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5 COT-EAT

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Cotton

A fiber obtained from the cotton plant *Gossypium*, of the order Malvales. It has been used for more than 3000 years. It is the most widely used natural fiber, because of its versatility and comparatively low cost to produce and to manufacture into finished products. Cotton traditionally has been used alone in textile products, but blends with artificial fibers have become increasingly important. *See* MALVALES; NATURAL FIBER; TEXTILE.

Unlike most other fibers obtained from plants, the cotton fiber is a single elongated cell. Under the microscope, it resembles a flattened, spirally twisted tube with a rough surface. The fiber cell has many convolutions, with a collapsed inner canal (lumen). Chemically, cotton is essentially pure cellulose. In its original state, cotton contains 3–5% natural waxes and gums, mostly in the outer wall of the cells. The natural waxes on the fiber surface act as a finish which facilitates spinning. Cotton is hygroscopic and contains 6–8% moisture under normal atmospheric conditions. *See* CELLULOSE.

Cultivation and harvesting. To mature, cotton requires about 180 days of continuously warm weather with adequate moisture and sunlight; frost is harmful and may kill the plant. The ground must be thoroughly plowed and the soil pulverized. In the United States, usually in March or April, carefully selected seeds are planted in rows. The plants require systematic fertilization. When they are about 3 in. (7.5 cm) high, they are weeded and thinned. The plants begin to bloom in June or July, about 4 months after planting. Creamy white flowers appear first, change to a reddish-purple in about 2 days, and then fall off, leaving seed pods that grow to full size by August or September. The cotton capsules or bolls must be protected against the boll weevil or other insects. *See* ENTOMOLOGY, ECONOMIC; FERTILIZER; FERTILIZING; FLOWER; FRUIT; SEED.

When fully grown, the cotton plant may be 3–6 ft (1–2 m) in height. Its wide green leaves partially conceal some of the bolls until they burst and expose the fleecy white fiber, which indicates that the cotton is ready for harvesting (**Fig. 1**). Not all cotton bolls open at the same time, but the ripening period has been shortened and pickings have been reduced to one or two. When the raw cotton is harvested, it contains seeds, leaf fragments, and dirt that must be removed before baling. The cotton seeds alone constitute approximately two-thirds of the weight of the picked cotton.

Products and processing. When the bolls open, the fiber and seed, or "seed cotton," is harvested mostly by machines. The fiber, or lint as it is then called, is separated from the seed by gins. The lint is compressed into bales, covered with jute bagging, and bound with metal bands for ease of handling. Bales weigh about 500 lb (225 kg) each. The seed (except

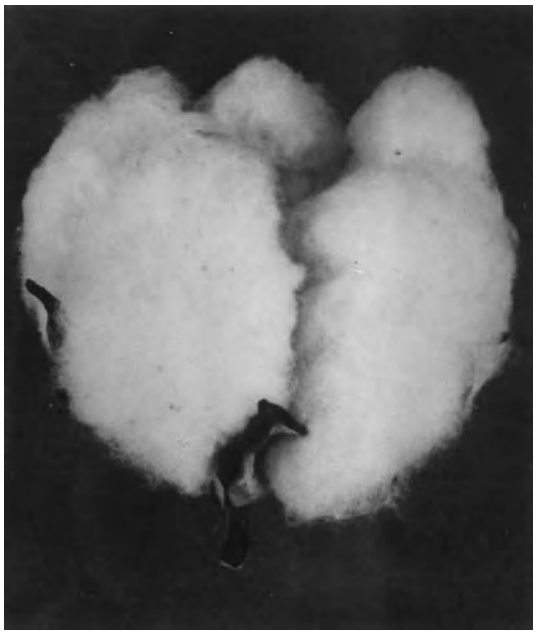

Fig. 1. Open cotton boll ready to harvest.

Fig. 2. Productive field of open cotton. (*K. Bilbrey, County Agent, Mississippi County, Arkansas*)

that portion needed for planting the next crop and for other farm uses) is transferred from the gins, usually by trucks, to oil mills for processing. After baling, the lint is sampled, graded, and sold. The bales then pass through commercial channels and eventually reach cotton mills, where they are broken open and the lint is blended, cleaned, carded, and spun into yarns for many uses. In the oil mill processing industry, the cottonseed is separated into fuzz or linters, hulls, oil, and protein cake. Each of these products is converted to several subproducts. The oil and cake are the most valuable products. The oil subproducts are largely used as human food; the protein cake, either as cracked cake or ground meal, is used as livestock feed. Hulls are also fed to livestock, and the linters are converted into chemical cellulose. *See* ANIMAL FEEDS; FAT AND OIL (FOOD).

Distribution and production. The cotton plant is one of the world's major agricultural crops. It is grown in all countries lying within a wide band around the Earth. The limits of this band in the New World are at about 37°N latitude and at about 32°S latitude. In the Old World the band spreads northward to 47°N in the Ukraine, but only to about 30°S in Africa and Australia. In addition to the effects of latitude, suitability of climate for the growth of cotton is also regulated by elevation, wind belts, ocean currents, and mountain ranges. As a result of the effects of these climatic factors, the topography of the land, the nature of the soil, and the availability of irrigation water when needed, cotton culture actually is carried out in an irregular and greatly extended world pattern. The regions of more intensive culture comprise the Cotton Belt of the United States, the northern valleys of Mexico, India, West Pakistan, eastern China, Central Asia, Australia, the Nile River Valley, East Africa, South Africa, northeastern and southern Brazil, northern Argentina, and Peru.

Stimulants to production. The United States has made three major contributions to world cotton production. These were the development of Upland cotton, the invention of the saw gin, and the development of knowledge of cotton culture.

Upland cotton arose from an annual form of *G. hirsutum* native to the plateaus of southern Mexico. This species was introduced into what is now southern United States at different times from the colonial period to the first decade of the twentieth century. The name Upland is derived from the upland country in southeastern United States, where this stock was first grown commercially. This cotton was shipped to England under the name of American Upland. From the time of its early development as a crop in the American Cotton Belt, it has been very hardy and productive, and versatile in its many uses (**Fig. 2**).

Upland production was handicapped at first by the difficulty of ginning. The seed coat has a tight covering of short fuzz hairs (**Fig. 3**). Only sawlike teeth can penetrate and remove the longer coat of lint from the seed and leave the fuzz behind with the seed coat undamaged. The saw gin, incorporating

Fig. 3. Ginned seed of Upland cotton, showing dense and tight covering of short fuzz hairs on seed coat. (*J. O. Ware, Journal of Heredity*)

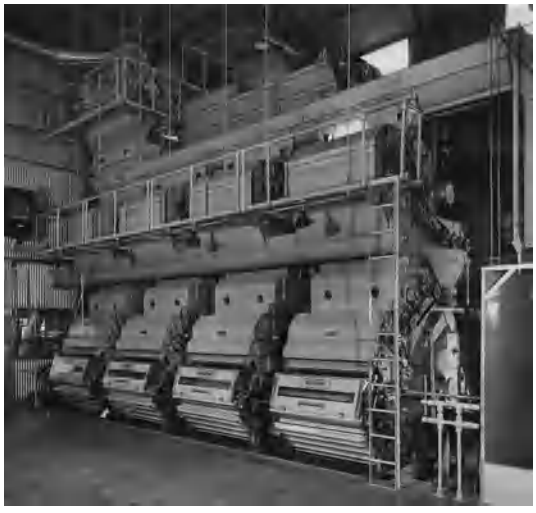

Fig. 4. Modern cotton gin, showing the four stands and the accessory equipment for cleaning and conveying the cotton. (*Murray Company of Texas*)

Fig. 5. Planting cotton with a four-row tractor-propelled planter. (*John Deere Co.*)

Fig. 6. Two-row, self-propelled cotton picker in operation. (*International Harvester Co.*)

this principle, was gradually developed into the modern gin plant (**Fig. 4**).

In the history of cotton culture, simple but efficient farm tools were developed and the mule was generally the source of farm power. The cotton fields were well tilled, the weeds controlled, the cotton carefully picked by hand, and the best seed selected for the next crop. However, when mechanical power began to replace animal power, the United States quickly mechanized both the cultural and harvesting phases of cotton production. (**Figs. 5** and **6**). Great progress in research, plant breeding, technology, and mechanics paved the way for advancements when economic conditions were ready for their adoption. Upland cotton and some of the improvements made in handling this crop spread to most cotton-growing countries as they began commercial production. Upland cotton provides about 85% of the world supply of raw cotton for factory use. *See* AGRICULTURAL SCIENCE (PLANT); BREEDING (PLANT).

The other 15% of world production is supplied from three other cultivated species: *G. barbadense, G. arboreum,* and *G. herbaceum*. The extra-long staple type includes Egyptian, Tanguis, American-Pima, and Sea Island. Egyptian is grown in Egypt and the Sudan, Tanguis in Peru, American-Pima in southwestern United States, and Sea Island in the West Indies. Sea Island cotton formerly was grown on islands and coastal areas of the southeastern United States. *Gossypium arboreum* and *G. herbaceum* are commonly referred to as Asiatic cottons and include the shortest-staple types of cultivated cottons. The growth of the former is largely confined to India and China and of the latter to Central Asia, western India, and the Near East. The saw gins usually used for Upland cotton in other countries are simpler and carry less cleaning equipment than those used in the United States. Most other countries still employ much handwork in cotton culture; most of the crop is harvested by hand.

Most commercial cottons are now annuals. However, in the tropics some races of the regular cultivated species are perennial. All cottons at one time were tropical and perennial.

Wild cotton. Besides the four cultivated species, *Gossypium* includes about 30 wild and lintless species that are widely scattered and occur mostly in sparsely covered desert and tropical areas of the world. All of the wild species may ultimately be of plant breeding interest. Some of them already have been used in crosses to improve quality values in cultivated cottons. Elton G. Nelson

Diseases. Pathogens may attack or invade the cotton plant through the roots or aboveground parts. Diseases reduce the yield potential 15 to 20%, and also have undesirable effects on quality of fiber and seed.

Root diseases are caused by soil-borne organisms. Severity of disease is proportional to the number of infective pathogen propagules in soil and to the degree of susceptibility in varieties. Minimizing losses involves using resistant varieties and practices that reduce pathogen numbers in the soil.

The seedling disease complex occurs wherever cotton is grown. Symptoms are seed rot, preemergence damping-off, postemergence damping-off, and damage to roots of plants that survive. The primary organisms are the fungi *Rhizoctonia solani, Pythium* spp., *Thielaviopsis basicola,* and *Glomerella gossypii* (**Fig. 7**), and the root-knot nematode *Meloidogyne incognita*.

Using high-quality seed treated with protectant fungicides is essential to the control of seedling

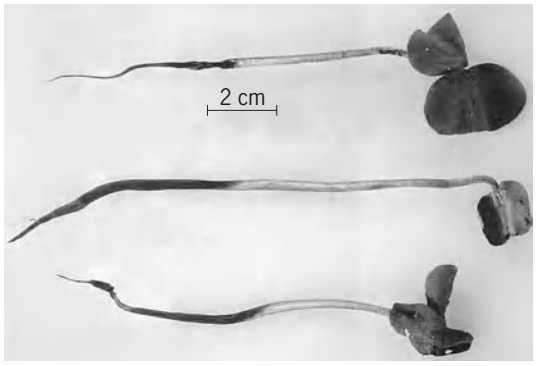

Fig. 7. Diseased cotton seedlings which have been affected by *Glomerella gossypii*.

Fig. 8. Verticillium wilt of young cotton plant.

diseases. Application of soil fungicides at planting is an additional practice that gives more effective control. Cotton varieties with resistance to seedling pathogens, preservation of seed quality, and with improved performance under early-season cool-wet conditions are now available.

Although the wilt-inducing fungus *Fusarium oxysporum* f. *vasinfectum* or the root-knot nematode *M. incognita* (Fig. 9*c*) can damage cotton in single infections, losses due to double infections are greater. Management practices effective against both organisms must therefore be used. The most economical practice is the use of resistant varieties, but it is best to use resistant varieties and rotations with root-knot–resistant crops such as small grains, corn, sorghum, or some grasses.

Verticillium wilt, caused by *Verticillium alboatrum*, occurs in the cooler regions of the world (**Fig. 8**). Effective control can be achieved with integrated management. High populations of 25 to 30 plants per meter of row help reduce losses. Too much nitrogen causes wilt to be more severe. Serious losses are usually limited to irrigated cotton, cotton in high-rainfall regions, and in poorly drained areas of fields. Reducing the amount of irrigation water by one-third will often give higher yields. The wilt fungus survives as microsclerotia in the soil and in undecomposed plant debris. Resistant crops such as barley and sorghum in rotation with cotton have been effective in reducing the inoculum potential. Many resistant varieties are available, and their use, along with other practices that reduce disease severity, gives effective control.

Root rot, caused by *Phymatotrichum omnivorum*, occurs in alkaline soils in the southwestern and western regions of the United States and in Mexico. The soil in infested areas has more sclerotia, lower bacterial populations, less organic matter, and less sodium than other soils. Control depends on reducing sclerotia production and preventing survival of the fungus between cropping seasons.

Burial of barnyard manure or residues of such legumes as Hubam clover and winter peas helps control the disease. Turning the upper 8–12 in. (20 to 30 cm) of soil with a moldboard plow a few days after harvesting a crop of cotton reduces sclerotia production. Rotation of cotton with grain crops also helps to keep numbers of sclerotia low.

Planting early in the season with fast-maturing varieties permits early boll set and maturation of the crop before soil temperatures reach levels for the pathogen to become highly active. Restoration of sodium to a level equal to that in noninfested soil prevents production of new sclerotia, and is now being used for control. Best control comes from using short-season management with practices that reduce sclerotia production and survival.

Diseases of aboveground plant parts may be caused by bacteria or fungi. These organisms may survive in undecomposed plant debris, on alternative hosts, or in and on seed. Sanitation of planting-seed and shredding and burial of residues from a cotton crop reduce the ability of foliage pathogens to survive the winter.

Bacterial blight, caused by *Xanthomonas malvacearum*, affects all plant parts and occurs wherever cotton is grown (**Fig. 9***a* and *b*). The bacterium survives in and on seed and as dried exudates on undecomposed plant debris. Chemical delinting

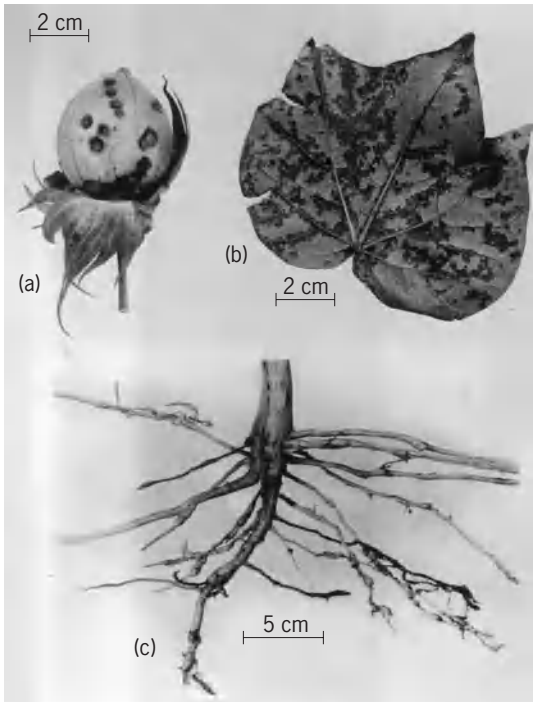

Fig. 9. Bacterial blight on cotton plant. (*a*) Bacterial blight lesions on bolls. (*b*) Bacterial blight lesions on leaves. (*c*) Root-knot galls on cotton roots.

removes the bacterium from the seed surface, but no treatment is available for eliminating internal transmission. Using seed from disease-free fields ensures no transmission. Management to ensure debris decomposition eliminates crop residue as a means of transmission. Varieties with high horizontal resistance to the 18 races of the pathogen are available for many cotton-growing regions of the world.

Boll rot may be caused by a number of fungi, bacteria, or yeasts. Some, such as *G. gossypii* and *X. malvacearum*, are primary pathogens. Many gain entrance through natural openings such as boll sutures and nectaries; others enter through wounds caused by insects or primary pathogens.

Control hinges on using practices that reduce inoculum densities and promote dryness within the leaf canopy. Varieties with okra-shaped leaves and frego bracts (which are strap-shaped leaves) are significant in reducing boll rot. Sanitation should be practiced to reduce transmission on seed and survival of organisms on debris. *See* PLANT PATHOLOGY.

Luther S. Bird

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Cotton effect

The characteristic wavelength dependence of the optical rotatory dispersion curve or the circular dichroism curve or both in the vicinity of an absorption band.

When an initially plane-polarized light wave traverses an optically active medium, two principal effects are manifested: a change from planar to elliptic polarization, and a rotation of the major axis of the ellipse through an angle relative to the initial direction of polarization. Both effects are wavelength dependent. The first effect is known as circular dichroism, and a plot of its wavelength (or frequency) dependence is referred to as a circular dichroism (CD) curve. The second effect is called optical rotation and, when plotted as a function of wavelength, is known as an optical rotatory dispersion (ORD) curve. In the vicinity of absorption bands, both curves take on characteristic shapes, and this behavior is known as the Cotton effect, which may be either positive or negative (**Fig. 1**). There is a Cotton effect associated with each absorption process, and hence a partial CD curve or partial ORD curve is associated with each particular absorption band or process. *See* OPTICAL ROTATORY DISPERSION; POLARIZED LIGHT.

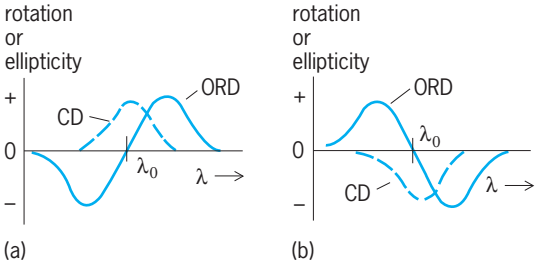

Fig. 1. Behavior of the ORD and CD curves in the vicinity of an absorption band at wavelength λ_0 (idealized). (*a*) Positive Cotton effect. (*b*) Negative Cotton effect.

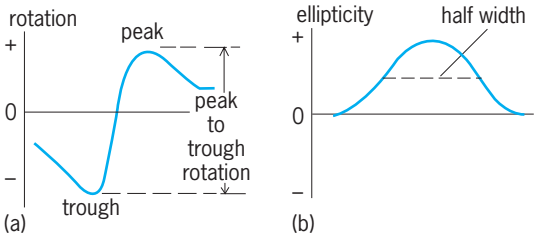

Fig. 2. Curves used to determine relative rotatory intensities. (*a*) Partial ORD curve. (*b*) Partial CD curve.

Measurements. Experimental results are commonly reported in either of two sets of units, termed specific and molar (or molecular). The specific rotation $[\alpha]$ is the rotation in degrees produced by a 1-decimeter path length of material containing 1 g/ml of optically active substance, and the specific ellipticity θ is the ellipticity in degrees for the same path length and same concentration. Molar rotation $[\varphi]$ (sometimes $[M]$) and molar ellipticity $[\theta]$ are defined by Eqs. (1) and (2). For comparisons among

$$[\varphi] = [\alpha]M/100 \tag{1}$$

$$[\theta] = \theta M/100 \tag{2}$$

different compounds, the molar quantities are more useful, since they allow direct comparison on a mole-for-mole basis.

The ratio of the area under the associated partial CD curve to the wavelength of the CD maximum is a measure of the rotatory intensity of the absorption process. Moreover, for bands appearing in roughly the same spectral region and having roughly the same half-width (**Fig. 2**), the peak-to-trough rotation of the partial ORD curve is roughly proportional to the wavelength-weighted area under the corresponding partial CD curve. In other words, relative rotatory intensities can be gaged from either the pertinent partial ORD curves or pertinent partial CD curves. A convenient quantitative measure of the rotatory intensity of an absorption process is the rotational strength. The rotational strength R_i of the ith transition, whose partial molar CD curve is $[\theta_i(\lambda)]$, is given by relation (3).

$$R_i \approx 6.96 \times 10^{-43} \int_0^\infty \frac{[\theta_i(\lambda)]}{\lambda} d\lambda \tag{3}$$

Molecular structure. The rotational strengths actually observed in practice vary over quite a few orders of magnitude, from $\sim 10^{-38}$ down to 10^{-42} cgs and less; this variation in magnitude is amenable to stereochemical interpretation. In this connection it is useful to classify optically active chromophores, which are necessarily dissymmetric, in terms of two limiting types: the inherently dissymmetric chromophore, and the inherently symmetric but dissymmetrically perturbed chromophore. *See* OPTICAL ACTIVITY.

A symmetric chromophore is one whose inherent geometry has sufficiently high symmetry so that the isolated chromophoric group is superimposable on its mirror image, for example, the carbonyl group $\rangle C{=}O$. The transitions of such a chromophore can become optically active, that is, exhibit a Cotton effect, only when placed in a dissymmetric molecular environment. Thus, in symmetrical formaldehyde, $H_2C{=}O$, the carbonyl transitions are optically inactive; in ketosteroids, where the extrachromophoric portion of the molecule is dissymmetrically disposed relative to the symmetry planes of the $\rangle C{=}O$ group, the transitions of the carbonyl group exhibit Cotton effects. In such instances the signed magnitude of the rotational strength will depend both upon the chemical nature of the extrachromophoric perturbing atoms and their geometry relative to that of the inherently symmetric chromophore. In a sense, the chromophore functions as a molecular probe for searching out the chemical dissymmetries in the extrachromophoric portion of the molecule.

The type of optical activity just described is associated with the presence of an asymmetric carbon (or other) atom in a molecule. The asymmetric atom serves notice to the effect that, if an inherently symmetric chromophore is present in the molecule, it is almost assuredly in a dissymmetric environment, and hence it may be anticipated that its erstwhile optically inactive transitions will exhibit Cotton effects. Moreover, the signed magnitude of the associated rotational strengths may be interpreted in terms of the stereochemistry of the extrachromophoric environment, as compared with that of the chromophore. But an asymmetric atom is not essential for the appearance of optical activity. The inherent geometry of the chromophore may be of sufficiently low symmetry so that the isolated chromophore itself is chiral, that is, not superimposable on its mirror image, for example, in hexahelicene.

In such instances the transitions of the chromophore can manifest optical activity even in the absence of a dissymmetric environment. In addition, it is very often true that the magnitudes of the rotational strengths associated with inherently dissymmetric chromophores will be one or more orders of magnitude greater ($\sim 10^{-38}$ cgs, as opposed to $< 10^{-39}$ cgs) than those associated with inherently symmetric chromophores. In the spectral regions of the transitions of the inherently dissymmetric chromophore, it will be the sense of handedness of the chromophore itself that will determine the sign of the rotational strength, rather than perturbations due to any dissymmetric environment in which the inherently dissymmetric chromophore may be situated.

The sense of handedness of an inherently dissymmetric chromophore may be of considerable significance in determining the absolute configuration or conformations of the entire molecule containing that chromophore. Accordingly, the absolute configuration or conformation can often be found by focusing attention solely on the handedness of the chromophore itself. For example, in the chiral molecule

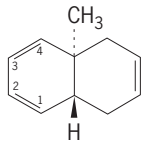

Fig. 3. Structural formula of (+)-*trans*-9-methyl-1,4,9,10-tetrahydronaphthalene.

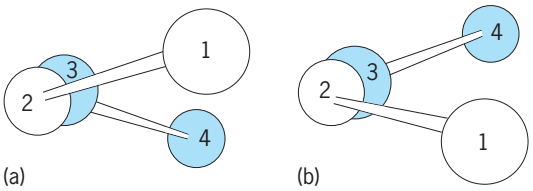

Fig. 4. Schematic representation of the twisted diene chromophore showing the two possible handednesses; the numbering is as indicated in Fig. 3. (*a*) Right-handed. (*b*) Left-handed.

shown in **Fig. 3** there is a one-to-one correspondence between the sense of helicity of the nonplanar diene chromophore present and the absolute configuration at the asymmetric carbon atoms. Hence there exists a one-to-one correspondence between the handedness of the diene and the absolute configuration of the molecule. Since it is known that a right-handed diene helix (**Fig. 4**) associates a positive rotational strength with the lowest diene singlet transition in the vicinity of 260 nanometers, by examination of the pertinent experimental Cotton effect (positive), the absolute configuration of the molecule is concluded to be as shown.

Other examples of inherently dissymmetric chromophores are provided by the helical secondary structures of proteins and polypeptides. Here the inherent dissymmetry of the chromophoric system arises through a coupling of the inherently symmetric monomers, which are held in a comparatively fixed dissymmetric disposition relative to each other through internal hydrogen bonding. The sense of helicity is then related to the signs of the rotational strengths of the coupled chromophoric system. The destruction of the hydrogen bonding destroys the ordered dissymmetric secondary structure, and there is a concomitant decrease in the magnitude of the observed rotational strengths. Albert Moscowitz

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Coulomb excitation

Nuclear excitation caused by the time-dependent long-ranged electric field acting between colliding nuclei. Theoretically, the Coulomb force between the positively charged colliding nuclei is well understood, and the interaction is exactly calculable. Coulomb excitation usually is the dominant reaction in nuclear scattering, and even occurs at low bombarding energies where the separation of

the nuclei is sufficiently large that the short-ranged nuclear force does not act. *See* COULOMB'S LAW.

Coulomb excitation plays a vital role in probing the response of both shape and volume collective modes of motion as well as the interplay of single-particle degrees of freedom of the nuclear many-body system. The goal of this work is to develop better models of nuclear structure and to elucidate the underlying nuclear force.

Collective nuclear modes of motion. The residual interaction between nucleons bound in the nucleus leads to coherent collective modes of motion of the nuclear surface and volume. Such coherent motion of many nucleons in the nucleus is of considerable interest in understanding the physics of many-body quantal systems. Collective rotation and vibration of deformed shapes of the nuclear surface is a dominant and ubiquitous feature of the low-lying structure in nuclei. Quite separate from these low-lying surface collective modes are volume collective modes that lead to high-frequency giant resonances at 10–30 MeV in excitation energy. The nuclear charge of the nucleons involved in this coherent motion produces considerably enhanced electromagnetic properties for collective nuclear states. The electric multipole moments of the nuclear states are a direct and sensitive measure of nuclear deformation. For example, the electric quadrupole moments are a direct measure of quadrupole deformation, such as football-shaped deformation; the electric octupole moments are sensitive to octupole deformation, such as pear shapes; while electric hexadecapole moments are sensitive to more complicated hexadecapole-shaped deformation. *See* GIANT NUCLEAR RESONANCES; NUCLEAR MOMENTS.

The considerable importance of Coulomb excitation lies in the fact that it is the preeminent probe of collective-shape degrees of freedom in nuclei. That is, Coulomb excitation selectively populates modes of motion of the collective shape with cross sections that are a direct and sensitive measure of the electric moments, and these electric moments can be measured with considerable precision since the electromagnetic interaction is exactly calculable.

One-step and multistep excitation. Coulomb excitation was first observed in the 1950s and played a pivotal role in showing that many nuclei have prolate deformation like a football. The initial experiments used beams of protons or alpha particles for which the electromagnetic interaction is weak and only simple one-step excitation occurs. The introduction of high-atomic-number (high-Z) projectiles dramatically advanced exploitation of Coulomb excitation. With such projectiles, the electromagnetic excitation probability for surface modes can approach unity and multistep excitation dominates, leading to the population of excited states with up to 34 units of angular momentum. Such multistep Coulomb excitation can determine the electromagnetic properties of many low-lying collective states in a nucleus, making it a powerful probe of collective motion in nuclear structure. *See* ALPHA PARTICLES; PROTON.

Technical advances. Three technical advances have greatly enhanced the power of Coulomb excitation as a probe of nuclear structure. The first is the development of heavy-ion accelerators that can provide copious beams of stable nuclear isotopes throughout the periodic table, including high-Z projectiles such as uranium ($Z = 92$). Also, radioactive beam facilities are being built that expand the arsenal of beams to include unstable nuclear species. The second advance is the fabrication of arrays of large intrinsic-germanium high-resolution gamma-ray detectors that surround the target. These have high detection efficiency and extremely high sensitivity for resolving the gamma rays emitted during the subsequent decay of excited states populated by multiple Coulomb excitation. In addition, large-solid-angle arrays of heavy-ion detectors are used to detect the coincident scattered ions and to determine unambiguously the nuclei excited and the scattering trajectories. The third advance is the development of a Coulomb excitation least-squares search computer code that makes it possible to extract the hundreds of electromagnetic matrix elements that couple the many states excited in multiple Coulomb excitation.

These advances allow population and study of complete sets of states for low-lying collective bands in a nucleus. Coulomb excitation has allowed detailed mapping of the collective-shape degrees of freedom in nuclei. The moments of inertia of collective rotation bands are derived from the excitation energies, while the nuclear shapes are derived from the electric moments. *See* GAMMA-RAY DETECTORS; PARTICLE ACCELERATOR.

Studies of low-lying modes. Coulomb excitation has produced a wealth of information on low-lying collective-shape degrees of freedom in nuclei. It has allowed study of rotational bands up to high angular momentum in nuclei that are not easily populated by other reaction mechanisms, such as neutron-rich stable nuclei, transuranic nuclei, and radioactive neutron-rich nuclei produced at radioactive beam facilities. Such studies have shown that collective motion is richer than early theoretical models had suggested. In many nuclei, Coulomb excitation has identified complete sets of states in rotational bands that result from rotation of football shapes with axis ratios of about 1.5 to 1. These strongly deformed prolate quadrupole shapes have electric quadrupole transition strengths that are over 200 times greater than those produced by a single proton. Other collective nuclear states have been found corresponding to the rotation and vibration of nearly oblate quadrupole deformed shapes, but where all three spatial axes of the nuclear shape differ in length. Collective states attributed to both one and two units of quadrupole or octupole vibration have been discovered. More complicated motion, such as bands of states corresponding to pear-shaped octupole vibration about rotating prolate deformed shapes, have been studied. Coexistence of rotational-vibrational bands having very different deformation also has been discovered in certain nuclei. These

observations are being used to refine models of nuclear structure.

Studies of giant resonances. The above studies involve Coulomb excitation of low-lying rotational and vibrational collective surface modes. Scattering of much faster heavy ions at very small scattering angles can lead to distances of closest approach that still are large enough to ensure that the interaction is dominated by the electromagnetic interaction. The shorter electromagnetic impulse in such fast Coulomb excitation makes it possible to excite the high-frequency collective volume modes, that is, giant resonances. Fast Coulomb excitation has been used to map giant resonances corresponding to both in-phase and out-of-phase vibrations of the protons and neutrons as well as double-phonon giant resonances. These studies elucidate the interplay of collective and single-particle degrees of freedom in nuclear structure. *See* NUCLEAR STRUCTURE; SCATTERING EXPERIMENTS (NUCLEI).

Douglas Cline

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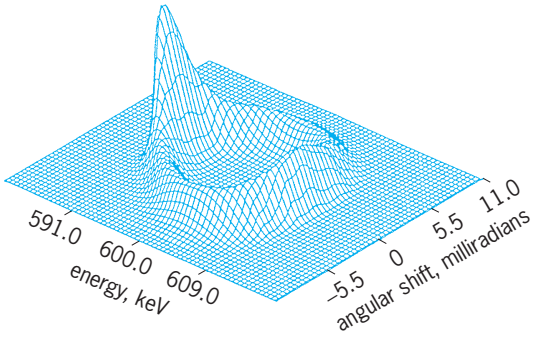
Coulomb explosion

A process in which a molecule moving with high velocity strikes a solid and the electrons that bond the molecule are torn off rapidly in violent collisions with the electrons of the solid; as a result, the molecule is suddenly transformed into a cluster of charged atomic constituents that then separate under the influence of their mutual Coulomb repulsion. The initial velocity of the molecule is typically greater than 3×10^6 ft/s (10^6 m/s), and it takes on the order of 10^{-17} s for electrons to be torn off the molecule. Typically, it takes about 10^{-15} s for the initial Coulomb potential energy of the cluster to be converted into kinetic energy as the charged fragments recede from one another. *See* COULOMB'S LAW.

Coulomb explosions are most commonly studied using a particle accelerator, normally employed in nuclear physics research (Van de Graaff generator, cyclotron, and so forth), to produce a beam of fast molecular ions that are directed onto a solid-foil target. The Coulomb explosion of the molecular projectiles begins within the first few tenths of a nanometer of penetration into the foil, continues during passage of the projectiles through the foil, and runs to completion after emergence of the projectiles into the vacuum downstream from the foil. Detectors located downstream make precise measurements of the energies and charges of the molecular fragments together with their angles of emission relative to the beam direction. The Coulomb explosion causes the fragment velocities to be shifted in both magnitude and direction from the beam velocity. The corresponding shifts in energy and angle are small, but if the foil target is thin (approximately 10 nm) and of light material (for example, carbon), the blurring effects of energy-loss straggling and multiple scattering in the foil can be kept small relative to the Coulomb explosion effects. *See* CHARGED PARTICLE BEAMS; PARTICLE ACCELERATOR.

Consider a beam of 3-MeV HeH^+ ions incident on a 10-nm-thick carbon foil. Upon striking the foil, each projectile produces a 2.4-MeV alpha particle and a 600-keV proton separated by about 80 picometers (the bond length for HeH^+). The Coulomb explosion causes the separation to grow to about 100 pm during traversal of the foil. Downstream, the fragments achieve asymptotic shifts in energy and angle that are determined by the initial orientation of the molecule. The maximum energy shift (± 8.4 keV) is obtained when the internuclear vector in the projectile is parallel to the beam direction. There is then no angular shift. The sign of the energy shift for a given fragment depends on whether it is leading or trailing its partner. If the internuclear vector is perpendicular to the beam, the maximum angular shift (0.4° for the proton, 0.1° for the alpha particle) is achieved, and there is no shift in energy. A joint distribution in energy and angle of protons from similar dissociations (with a 19.5-nm-thick foil) is shown in the **illustration**. The width of the "rim" reflects variations in the internuclear distance of the incident molecular ions due to their vibrational motion.

Coulomb explosion experiments of this type serve two main purposes. First, they yield valuable information on the interactions of fast ions with solids. For example, it is known that a fast ion generates a polarization wake that trails behind it as it traverses a solid. This wake can be studied in detail by using diatomic molecular-ion beams, since the motion of a trailing fragment is influenced not only by the Coulomb explosion but also by the wake of its partner (rather like the way a boat is affected by the wake of another in front of it). The nonuniform distribution of intensity around the ring in the illustration can be accounted for quantitatively in terms of forces due to the polarization wakes. Second, Coulomb-explosion



Joint distribution in energy and angle for protons from the dissociation of 3-MeV HeH^+ ions incident on a 19.5-nm-thick carbon foil.

techniques can be used to determine the stereochemical structures of molecular-ion projectiles. For example, with this method it was demonstrated experimentally for the first time that the H_3^+ molecule is equilateral-triangular. *See* ELECTRON WAKE; MOLECULAR STRUCTURE AND SPECTRA; STEREOCHEMISTRY.

Donald S. Gemmell

Coulomb's law

For electrostatics, Coulomb's law states that the direct force F of point charge q_1 on point charge q_2, when the charges are separated by a distance r in free space, is given by $F = k_0 q_1 q_2 / r^2$, where k_0 is a constant of proportionality whose value depends on the units used for measuring F, q, and r. It is the basic quantitative law of electrostatics. In the International System (SI) of units, $k_0 = 1/(4\pi\epsilon_0)$, where ϵ_0 is called the permittivity of empty space and has the value 8.85×10^{-12} farad/m. Thus, Coulomb's law is as in the equation below where q_1 and q_2 are expressed

$$F = \frac{1}{4\pi\epsilon_0}\frac{q_1 q_2}{r^2}$$

in coulombs, r is expressed in meters, and F is given in newtons. *See* ELECTRICAL UNITS AND STANDARDS.

The direction of F is along the line of centers of the point charges q_1 and q_2. The force is one of attraction if the charges are opposite in sign and one of repulsion if the charges have the same sign. For a statement of Coulomb's law as applied to point magnet poles. *See* MAGNETOSTATICS.

Experiments have shown that the exponent of r in the equation is very accurately the number 2. Ernest (Lord) Rutherford's experiments, in which he scattered alpha particles by atomic nuclei, showed that the equation is valid for charged particles of nuclear dimensions down to separations of about 10^{-12} cm. Nuclear experiments have shown that the forces between charged particles do not obey the equation for separations smaller than this.

The direct force that one charged particle exerts on another is unaffected by the presence of additional charge and, in any electrostatic system, the equation gives this direct force between q_1 and q_2 under any conditions of charge configuration, including that in which intervening and surrounding matter is present and the molecules of the matter are polarized so that their charges contribute to this configuration. The total force on any one charge, say q_1, is the vector sum of the separate direct forces on q_1 due to q_2, q_3, q_4, and so on, each force computed separately by use of the equation as if all other charges were absent.

The permittivity ϵ of a medium is defined by $\epsilon = \epsilon_r \epsilon_0$, where ϵ_r is the relative permittivity of the medium. It was formerly known also as the relative dielectric constant or specific inductive capacity. *See* PERMITTIVITY.

If two free point charges q_1 and q_2 are immersed in an infinite homogeneous isotropic dielectric, the total force on one of them, say q_1, is given by $F = q_1 q_2/(4\pi\epsilon r^2)$ and the use of ϵ (in place of ϵ_0) takes proper account of the forces on q_1 due to the polarization charges of the dielectric molecules. *See* ELECTRIC CHARGE; ELECTROSTATICS.

Ralph P. Winch

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Coulometer

Electrolysis cell in which a product is obtained with 100% efficiency as a result of an electrochemical reaction. The quantity of electricity, that is, the number of coulombs of electricity (Q), can be determined very accurately by weighing the product that is deposited on an electrode in the course of the electrochemical reaction. The relationship between the weight of the product formed in the coulometer and the quantity of electricity used is given by Faraday's laws of electrolysis. When a constant current of i amperes flows through the electrolyte in the coulometer for t seconds, the number of coulombs passed is given by Eq. (1). If the current varies in the course of the

$$Q = it \tag{1}$$

electrolysis, the simple current-time product in Eq. (1) is replaced by the current-time integral, Eq. (2). When Q coulombs of electricity are passed

$$Q = \int_0^t i\,dt \tag{2}$$

through the electrolyte, the weight in grams of the material that is deposited on the electrode (w) is given by Eq. (3), where n is the number of electrons

$$w = \frac{QM}{Fn} \tag{3}$$

transferred per mole of material deposited, M is its molecular weight, and F is the Faraday constant, 96,487 ± 1.6 coulombs.

Equation (3) is fundamental in coulometry and is a mathematical statement of Faraday's laws. This equation is used for the accurate determination of Q, the current-time integral, by weighing or measuring a product that is formed at an electrode by an electrochemical reaction that occurs with 100% current efficiency. The electrolysis cell that is used for this purpose is a coulometer.

Only a few electrode reactions proceed with the 100% current efficiency that is required for the use of Eq. (3). The deposition of silver or copper (in a silver or copper coulometer), the evolution of oxygen and hydrogen (in a gas coulometer), and the oxidation of iodide to iodine (in an iodine coulometer) are examples of electrode reactions that have been successfully employed. One coulomb of electricity will deposit 1.1180 mg of silver at the cathode in

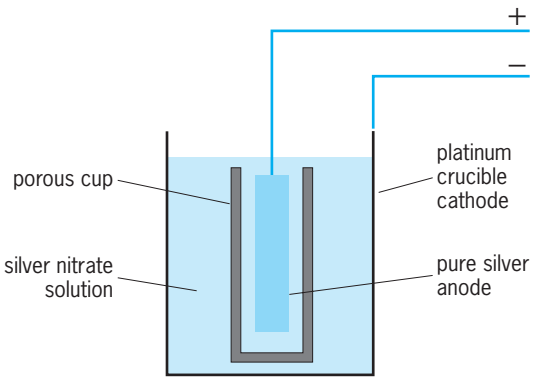

Fig. 1. Silver coulometer. The porous cup catches particles of silver that fall off the anode.

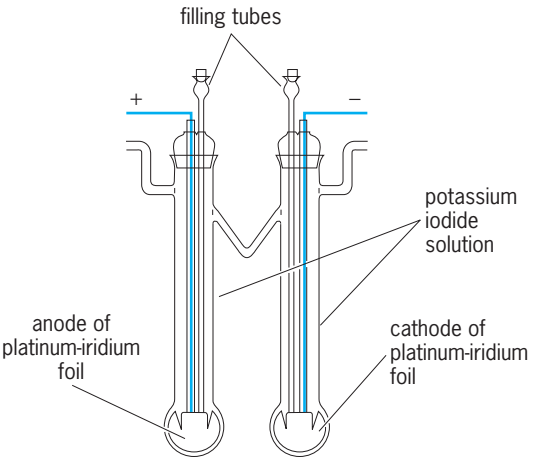

Fig. 2. Iodine coulometer.

a silver coulometer (**Fig. 1**) or liberate 1.315 mg of iodine at the anode in an iodine coulometer (**Fig. 2**). Although these classical chemical coulometers are capable of measuring the quantity of electricity with high precision and accuracy, their use is time-consuming and inconvenient; and they have been largely replaced by operational amplifier integrator circuits or digital circuits that display in a direct readout the number of coulombs passed during electrolysis. *See* ELECTROCHEMICAL EQUIVALENT; ELECTROLYSIS. Quintus Fernando

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Countercurrent exchange (biology)

Engineers have known for decades that efficient, almost complete heat or other exchange could be achieved between two fluids flowing in opposite directions in separate tubes. Such countercurrent systems have been "invented" numerous times by living organisms for all types of exchange function. They are most commonly found in the circulatory, respiratory, and excretory (kidney) systems, serving in heat, oxygen, and ion exchange. Biological countercurrent systems can be classified into two main types: downhill exchanges and hairpin multipliers. In both cases, the basic mechanism is the same—exchange of substance between fluids flowing in opposite directions—but the consequences are very different.

Downhill exchanges. These countercurrent systems are commonest in the circulatory system where their morphological structure is a rete mirabile (a wonderful net) of closely apposed sets of small arteries and veins. They are also found in gills of fish and in the minute air tubules of the avian lung. The principle of downhill exchanges is simple, as shown in **Fig. 1**. Fluids flow in opposite directions in separate tubes with the possibility of exchange, for example, heat flow or diffusion of oxygen, between them. The fluid entering one tube is warmest at that end, while that entering the second tube is coolest at the other end. Heat flows from higher to lower temperature. As heat flows from the warmer to the cooler tube, the fluid in the warmer tube cools down slightly and moves down along the tube. But as the slightly cooler fluid has moved further, it comes into contact with still cooler fluid in the second tube, and additional heat flow can occur. Thus, as the warmer fluid flows down the tube, it constantly loses heat, but always comes into contact with even cooler fluid in the second tube; the reverse is true for the initially cool fluid as it receives heat and warms up. Although the temperature differential between the two fluids is small at any point along the length of the countercurrent system, almost all the heat contained in the warmer tube is transferred to the cooler tube. Exchange of heat or oxygen occurs by passive diffusion. Most of the heat that entered the countercurrent system at one end leaves the system at the same end.

Retia of blood vessels thus serve as thermal isolating mechanisms within the body. They are found in appendages of mammals and birds (for example,

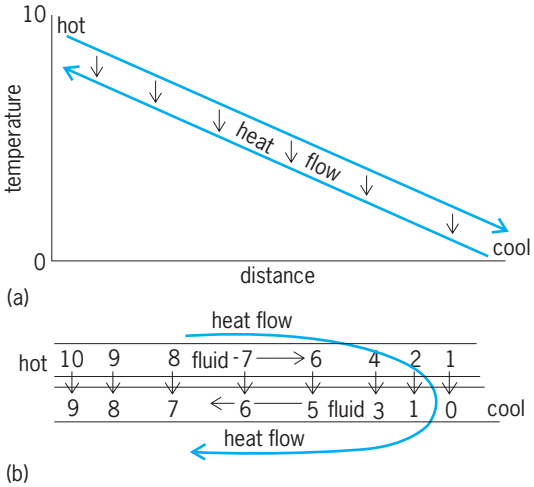

Fig. 1. Downhill exchange countercurrent system. (*a*) Graph showing the relationship between distance and temperature in a downhill system. (*b*) Schematic showing the two tubes carrying hot fluid and cool fluid and the mechanism by which heat exchange takes place.

whale flippers, the tail of beavers, and legs of gulls) to prevent excessive heat loss from these uninsulated parts. In reverse, masses of warm muscles in rapidly swimming fish, such as mackerel, tuna, and the mako shark, are isolated from the rest of the body and the gills, where heat loss would occur, by sets of retia. In desert mammals a rete located between the veins draining evaporating (that is, cooling) surfaces and the carotid artery cools the arterial blood before it reaches the brain.

Downhill exchange systems in the gills of fish and in the air tubules of birds permit maximum exchange of oxygen from the environment into the blood. Blood in respiratory capillaries flows against the water or air current and thus can pick up most of the oxygen contained in the external fluid. The advantage of downhill exchangers is that they achieve greater efficiency without extra energy cost simply by arranging flow in a countercurrent rather than in a concurrent fashion.

Hairpin multipliers. These exchange systems take their name from the structure of the tubes, which have a hairpin turn between the afferent (descending) and the efferent (ascending) limbs. Hairpin countercurrent systems are found in the nephron (the loop of Henle) of the kidney and in the capillary system of the gas gland in the swim bladder of many fish. In contrast to downhill systems, which operate by passive transport, hairpin multipliers must employ active transport of materials. These are always materials pumped out of the efferent limb of the system.

Operation of a hairpin multiplier, such as the loop of Henle, is shown in **Fig. 2**. As the kidney filtrate flows down the afferent limb and up the efferent limb, sodium ions (Na^+) are actively transported out of the efferent limb. The sodium diffuses back into the afferent limb and is carried once again around the hairpin turn, together with additional sodium that constantly enters the system via the afferent limb. Continual active transport of sodium along the entire length of the efferent limb and its diffusion back into the afferent limb will result in the accumulation of sodium at the bottom of the loop and in the surrounding interstitial fluid, and in the establishment of a steep osmotic gradient from the top to the bottom of the hairpin system. Yet at no point within this multiplier mechanism are individual cells exposed to excessive osmotic pressure. The steep gradient of osmotic pressure is used for final concentration of the urine as it passes down the collecting ducts. As water is drawn from the urine by the osmotic pressure of the interstitial fluid, the more concentrated urine passes into an area of even greater osmotic pressure—an example of downhill exchange. Thus, both types of countercurrent systems operate in the kidney to recover water from the urine. Desert mammals improve their water-conserving abilities by simply increasing the length of the loop of Henle; hence the strength of the osmotic pressure at the lower end of the hairpin is also increased.

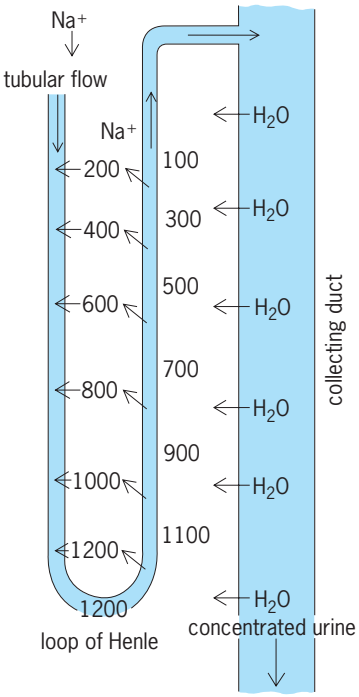

Fig. 2. Hairpin multiplier countercurrent system (loop of Henle of nephron).

The hairpin multiplier in the gas gland of the swim bladder of deep-sea marine fish serves to concentrate oxygen. Lactic acid is produced in the gas gland located at the bend of the hairpin capillaries and secreted into the blood. Lactic acid drives oxygen from hemoglobin faster than it can recombine with the hemoglobin molecule. The oxygen diffuses out of the efferent limb and into the afferent limb. With the constant addition of oxygen by arterial blood in the efferent limb, a very steep concentration gradient increase of over 1000-fold is achieved, thereby filling the swim bladder with gaseous oxygen against the great water pressures at depth of 330 ft (1000 m) or more. *See* KIDNEY; RESPIRATORY SYSTEM; SWIM BLADDER.

Walter J. Bock

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Countercurrent transfer operations

Industrial processes in chemical engineering or laboratory operations in which heat or mass or both are transferred from one fluid to another, with the fluids moving continuously in very nearly steady state or constant manner and in opposite directions through the unit. Other geometrical arrangements for transfer operations are the parallel or concurrent flow, where the two fluids enter at the same end of the apparatus and flow in the same direction to the other end, and the cross-flow apparatus, where the two fluids flow at right angles to each other through the apparatus. These two arrangements are ordinarily not as efficient as countercurrent flow, but do

find certain applications in industry and the laboratory.

Heat transfer. In heat transfer there can be almost complete transfer in countercurrent operations. The limit is reached when the temperature of the colder fluid becomes equal to that of the hotter fluid at some point in the apparatus. At this condition the heat transfer is zero between the two fluids. The amount of actual heat transfer is determined by economical design, that is, by comparing the value of the transferred heat with the cost of the heat exchange equipment. The economically optimum heat transfer has been studied for many years in engineering and is changing constantly as the costs of basic forms of energy increase.

Most heat transfer equipment has a solid wall between the hot fluid and the cold fluid, so the fluids do not mix. Heat is transferred from the hot fluid through the wall into the cold fluid. Another type of equipment, fewer in number but significant in size and use, does use direct contact between the two fluids—for example, the cooling towers used to remove heat from a circulating water stream. Cooling towers are of the countercurrent type and the cross-flow type. For more discussion of heat transfer *see* COOLING TOWER; HEAT BALANCE; HEAT EXCHANGER; HEAT TRANSFER.

Mass transfer. This process involves the changing compositions of mixtures, and is done usually by physical means instead of chemical reactions because of the lower costs. Even as heat is transferred from a region of high temperature to one of lower temperature, a material is transferred within a single phase from a region of high concentration to one of lower concentration by processes of molecular diffusion and eddy diffusion. In typical mass transfer processes, at least two phases are in direct contact in some state of dispersion, and mass (of one or more substances) is transferred from one phase across the interface into the second phase. Similar to heat transfer, mass transfer takes place between two immiscible phases until equilibrium between the two phases is attained. In heat transfer, equilibrium denotes an equality of temperature in the two phases, but in mass transfer there is seldom an equality of concentration in the two equilibrium phases. This means that a component may be transferred from a phase at low concentration (but at a concentration higher than that at equilibrium) to a second phase of greater concentration. The approach to equilibrium is controlled by diffusion transport across phase boundaries. Because this is a relatively slow process, the transport rate is increased by increasing the total interfacial area, by decreasing the thickness of the near-stagnant films adjacent to the interface, and by more frequent renewals of the interfacial films.

Although the two phases may be in concurrent flow or cross-flow, usual arrangements have the phases moving in countercurrent directions. The more dense phase enters near the top of a vertical cylinder and moves downward under the influence of gravity. The less dense phase enters near the bottom of the cylinder and moves upward under the influence of a small pressure gradient. In some cases, centrifugal force is used instead of gravity to provide phase separations.

For discussions of specific mass transfer operations *see* ADSORPTION; CHEMICAL SEPARATION TECHNIQUES; CRYSTALLIZATION; DISTILLATION; DRYING; ELECTROPHORESIS; EXTRACTION; LEACHING.

Frank J. Lockhart

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Couple

A system of two parallel forces of equal magnitude and opposite sense (**Fig. 1**). Forces P at normal distance P constitute a counterclockwise couple C_z (viewed from $+Z$) in the OXY plane; forces Q at arm q constitute a clockwise couple C_x (from $+X$) OYZ.

Under a couple's action a rigid body tends only to rotate about a line normal to the couple's plane. This tendency reflects the vector properties of a couple. *See* RIGID-BODY DYNAMICS.

Vector properties. The total force of a couple is zero. The total moment **C** of a couple is identical about any point. Accordingly, **C** is the moment of either force about a point on the other and is perpendicular to the couple's plane. *See* RESULTANT OF FORCES; STATICS.

In Fig. 1, $\mathbf{C}_z$, at the origin for convenience, is the moment of couple C_z. Its magnitude $|C_z| = +P_p$. Its sense, by the convention of moment, is $+Z$. Also $\mathbf{C}_x$, of sense $-X$, represents couple C_x; $|C_x| = +Q_q$.

Scalar moment. The moment of a couple about a directed line is the component of its total moment in the line's direction. For example, the moment of couple C_z about line L is $M_L = |C_z| \cos \theta_L = +P_p \cos \theta_L$. Also $M_x = 0$, $M_y = 0$, and $M_z = +P_p \cos 0^\circ = +P_p$.

Equivalent couples. Couples are equivalent whose total moments are equal. In **Fig. 2**, the paired linear forces and the counterclockwise curl represent

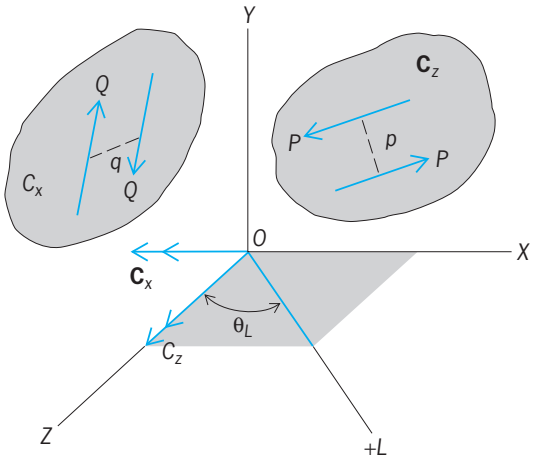

Fig. 1. Vector properties of a couple.

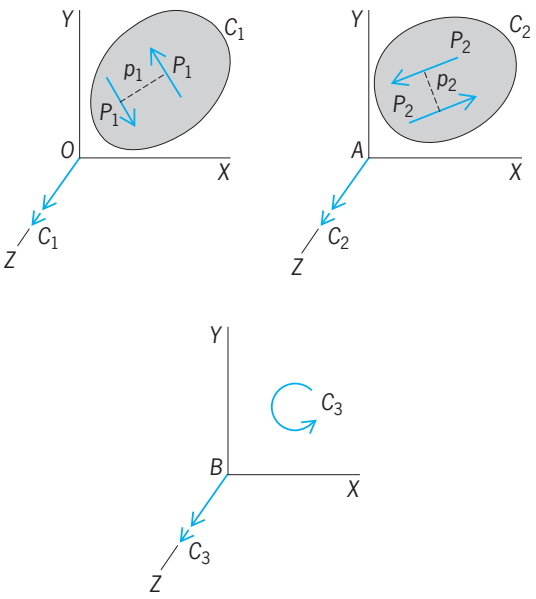

Fig. 2. Equivalent couples.

counterclockwise couple C_1, C_2, and C_3 in parallel planes OXY, AXY, and BXY. When their total moments are directed the same $(+Z)$, and their magnitudes $|C_1| = P_1p_1$, $|C_2| = P_2p_2$, and $|C_3|$ are equal, then $\mathbf{C}_1 = \mathbf{C}_2 = \mathbf{C}_3$ and these couples are equivalent. Thus $\mathbf{C}_1$ can represent these or any number of other equivalent couples. Nelson S. Fisk

Coupled circuits

Two or more electric circuits are said to be coupled if energy can transfer electrically or magnetically from one to another. If electric charge, or current, or rate of change of current in one circuit produces electromotive force or affects the voltage between nodes in another circuit, the two circuits are coupled.

Between coupled circuits there is mutual inductance, resistance, or capacitance, or some combination of these. The concept of a mutual parameter is based on the loop method of analysis. A mutual parameter can be one that carries two or more loop currents; such a network has conductive coupling because electricity can flow from one circuit to the other. *See* NETWORK THEORY.

Also, there can be purely inductive coupling, which appears if the magnetic field produced by current in one circuit links the other circuit. A two-winding transformer is an application of inductive coupling, with energy transferred through the magnetic field only.

It is also possible to have mutual capacitance, with energy transferred through the electric field only. An example is the capacitive interference between two transmission lines, as a power line and a telephone line, that run for a considerable distance side by side.

Polarity. With inductive coupling, polarity may need to be known, particularly if two circuits are coupled in more than one way. Do two kinds of coupling produce voltages that add or subtract?

There are several ways to show the relative polarities of inductive coupling. **Figure 1** shows, somewhat pictorially, two coils wound on the same core. Current flowing into the upper end of coil 1 would produce magnetic flux upward in the core, and so also would current flowing into the upper end of coil 2. For this reason the upper ends of the two coils are said to be corresponding ends.

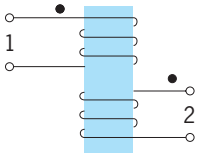

Fig. 1. Two coils wound on a single core, illustrating polarity.

If a wiring diagram is drawn with some such semipictorial sketch of the coils, it is not difficult to determine which are corresponding ends. However it is easier, both in representation and in interpretation, to indicate the corresponding ends symbolically. For this purpose dots are placed on a diagram at the corresponding ends of coupled coils. Such dots are shown in Fig. 1, though they are not needed there. Dots are also shown in **Fig. 2**, where they give the only means of identifying corresponding ends of the coils.

Note that the bottom ends of the coils shown in Fig. 1 are also corresponding ends, and that the two lower ends might have been dotted instead of the two upper ends; it makes no difference. But if the sense of winding of either coil 1 or coil 2 in Fig. 1 were reversed, then one dot (either one) would have to be shifted correspondingly.

Voltage equations. In addition to showing corresponding ends of two coils, Fig. 2 indicates that the two coils couple to each other. This coupling is identified as L_{12}. There is voltage from a to b in Fig. 2 if current i_1 is changing through the self-inductance L_{11} of the coil in circuit 1. There is additional voltage from a to b if current i_2 is changing in circuit 2 because of the mutual inductance L_{12} between the circuits. Thus, with circuit 1 coupled to circuit 2 and to any number of other circuits, Eqs. (1) and (2) hold. Here

$$v_{ab} = \left(L_{11}\frac{di_1}{dt} \pm L_{12}\frac{di_2}{dt} \pm \cdots\right) \qquad (1)$$

$$v_{cd} = \left(\pm L_{21}\frac{di_1}{dt} + L_{22}\frac{di_2}{dt} \pm \cdots\right) \qquad (2)$$

L_{11} and L_{22}, the self-inductances of the circuits, are positive numbers (inductances measured in henrys).

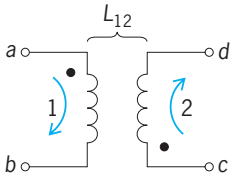

Fig. 2. Mutual inductance. Dots identify corresponding ends of coils.

With regard to the mutual terms, two questions of polarity must be asked. Are the mutual inductances such as L_{12} and L_{21} always positive numbers? And are the signs before the mutual terms, shown as $\pm$ in Eq. (1), actually $+$ or are they $-$? Answers to these questions are not always the same, and may be different as given by different authors, but the most usual simple procedure to answer them is as follows:

1. Draw a circuit diagram, such as Fig. 2, with nominal positive directions of currents shown by arrows and with dots at corresponding ends of coupled coils.

2. Let each mutual inductance such as L_{12} be a positive number.

3. Write equations such as Eqs. (1) and (2) with the following signs: First, if both arrows enter dotted ends of a pair of coupled coils, or if both arrows enter undotted ends, use $+$ before the corresponding mutual-impedance term. Second, if one arrow enters a dotted end and the other an undotted end, use $-$ before the corresponding mutual-impedance term.

However, this simple procedure sometimes fails, for it may be impossible to dot corresponding ends of all pairs of coupled coils in a network if there are three or more coils. A more general method follows:

1. Draw a circuit diagram with nominal positive directions of currents shown by arrows; place dots at coil ends arbitrarily, as may be convenient.

2. Determine corresponding ends of pairs of coupled coils, considering the coils two at a time. For the mutual inductance between a pair of coils with corresponding ends dotted, use a positive number, such as $L_{12} = 5$. For mutual inductance between a pair of coils with noncorresponding ends dotted, use a negative number, such as $L_{12} = -5$.

3. Write equations according to rule 3, above.

Steady-state equations. The differential Eqs. (1) and (2) are quite general. For steady-state operation at a single frequency, it is often simpler to have phasor or transform equations. Such equations can be in terms of reactances instead of inductances ($X = \omega L = 2\pi fL$ where f is frequency in cycles per second, or hertz). With the usual interpretation of phasor or transform equations, the steady-state relations corresponding to Eqs. (1) and (2) are shown in Eqs. (3) and (4). The rules in the foregoing paragraphs deter-

$$\begin{aligned} V_{ab} &= j\omega(L_{11}I_1 \pm L_{12}I_2 \pm \cdots) \\ &= j(X_{11}I_1 \pm X_{12}I_2 \pm \cdots) \end{aligned} \tag{3}$$

$$\begin{aligned} V_{cd} &= j\omega(\pm L_{21}I_1 + L_{22}I_2 \pm \cdots) \\ &= j(\pm X_{21}I_1 + X_{22}I_2 \pm \cdots) \end{aligned} \tag{4}$$

mine the choice of $+$ or $-$ in these equations; also, a mutual reactance such as X_{12} is a positive number (of ohms) if the corresponding mutual inductance such as L_{12} is a positive number (of henrys), as determined by the above rules. *See* ALTERNATING-CURRENT CIRCUIT THEORY.

Equality of mutual inductance. Because of this equality it is not uncommon, if there are only two coupled coils in a network, to use the letter M in place of both L_{12} and L_{21}. This use of M will be adopted in the following paragraphs.

Coefficient of coupling. The coefficient of coupling between two circuits is, by definition, Eq. (5). If the

$$k = M/\sqrt{L_{11}L_{22}} \tag{5}$$

circuits are far apart or, because of orientation, have little mutual magnetic flux, they are loosely coupled and k is a small number. Values of k for circuits with loose coupling may be in the range between 0.01 and 0.10. For closely (or tightly) coupled circuits with air-core coils, k may be around 0.5. In a transformer with a ferromagnetic core, k is very nearly 1.00.

Ideal transformers. An ideal transformer is defined as one in which primary and secondary currents are related inversely as the number of turns in the windings; this is shown in Eq. (6). Voltages across the

$$I_1/I_2 = N_2/N_1 \tag{6}$$

primary and secondary windings are in direct proportion to the numbers of turns, as shown in Eq. (7).

$$V_1/V_2 = N_1/N_2 \tag{7}$$

An actual transformer may be almost ideal or not at all ideal, depending on how it is made, and construction in turn depends on the purpose for which it is to be used. In an ideal transformer $k = 1$; in any actual transformer k is less than 1. In an actual transformer there is magnetizing current, so Eq. (6) is not exact. Also, an actual transformer has resistance and leakage reactance, so Eq. (7) is not exact. Nevertheless, the relations of ideal transformers are closely approximated by 60-Hz power transformers, and these relations are more or less close for other transformers that have ferromagnetic cores. *See* TRANSFORMER.

Whereas coupling k is desirably as close as possible to unity in a transformer that couples power from one circuit to another, k may purposefully be considerably less than unity in a transformer used for another purpose. In an oscillator, coupling need only be sufficient to sustain oscillation. In a band-pass amplifier, coupling is determined by bandwidth requirements. *See* OSCILLATOR; RESONANCE (ALTERNATING-CURRENT CIRCUITS).

Equivalent circuits. It is often convenient to substitute into a network, in place of a pair of inductively coupled circuits, an equivalent pair of conductively coupled circuits. Circuits so substituted are equivalent if the network exterior to the coupled circuits is unaffected by the change; in many cases this requirement implies that input current and voltage and output current and voltage are unaffected by the change.

Voltages and currents at the terminals of the coupled circuits of **Fig. 3** are related by Eqs. (8) and (9).

$$(R_1 + j\omega L_{11})I_1 - j\omega MI_2 = V_{ab} \tag{8}$$

$$(R_2 + j\omega L_{22})I_2 - j\omega MI_1 = V_{cd} \tag{9}$$

Voltages and currents at the terminals of the network of **Fig. 4** are also related by Eqs. (8) and (9), so the

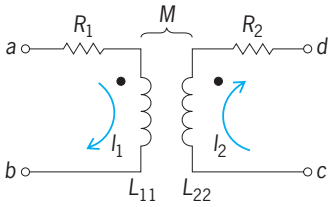

Fig. 3. A pair of coupled circuits.

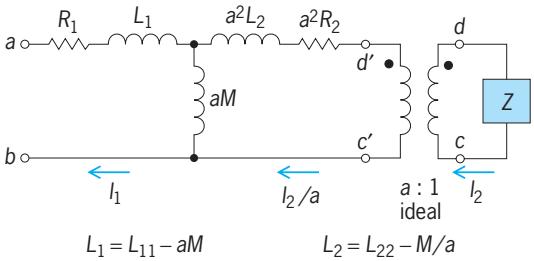

Fig. 4. A network equivalent to coupled circuits, feeding a load of impedance *Z*.

network of Fig. 4 is equivalent to the coupled circuits of Fig. 3. This is not immediately obvious, but it appears when loop equations for the network are simplified to a form that shows the requirement that $L_{11} = L_1 + aM$ and $L_{22} = L_2 + M/a$ and that $V_{c'd'} = aV_{cd}$, where a is the turn ratio of the ideal transformer shown, equal to N_1/N_2 in Eqs. (6) and (7).

It seems at first that this substitution, resulting in the network of Fig. 4, has produced something more complicated than the circuit of Fig. 3, but the value of substituting will now be shown.

Mathematically, a could be any number, but practically there are two particularly advantageous values for a. When the coupling between coils is loose so that k, the coefficient of coupling, is small, and if the coils have somewhere near the same number of turns, it is advantageous to let $a = 1$. With $a = 1$ the network of Fig. 4 is simplified to **Fig. 5**, leaving only the conductively coupled circuits with parameters R_1 and L_1, M, and L_2 and R_2, where $L_1 = L_{11} - M$, and $L_2 = L_{22} - M$. Many loosely coupled circuits particularly in radio circuits, are conveniently represented by this equivalent network with $a = 1$.

On the other hand, if coupling is close and especially when the two coupled coils have widely different numbers of turns (a situation that is typical of transformers), it is more convenient to let a equal the actual turn ratio of the coils. If, for example, there are 10 times as many turns in the primary winding of a transformer as there are in the secondary winding, it is well to let $a = 10$. (Letting $a = 1$ in such a transformer would result in a negative value for L_2 which, though correct for analysis, is difficult to visualize.)

With a equal to the actual transformer turn ratio, the following interpretation of Fig. 4 is usual and convenient. All causes of power loss and voltage drop have been put into the equivalent T network. Only the turn ratio, the actual transforming function, remains in the ideal transformer. It now becomes possible to consider, to study, and even to design the separate functions independently.

Transformers. The equivalent circuit of Fig. 4 is so commonly used in transformer work that a number of the quantities have been given special names. With the concept that power is supplied to one winding of the transformer, this is called the primary winding; the other, the secondary winding, provides power to a load. There may be a third, or tertiary, winding, perhaps providing power to another load at a different voltage or with a different connection, and even other windings, on the same transformer core.

However, it is only when a transformer has become part of a system that there is any meaning in designating the windings as primary and secondary, for a two-winding transformer can be used to carry power in either direction. The terms "high-voltage side" and "low-voltage side" are preferable for a transformer that is not part of a system. In the following discussion the words primary and secondary really mean nothing more than the windings of the transformer that are designated by the subscripts 1 and 2.

Speaking of a transformer with N_1 primary turns and N_2 secondary turns, let a be the turn ratio N_1/N_2. Then, referring to Fig. 4, R_1 is primary resistance; L_1 is primary leakage inductance; aM is primary exciting inductance; a^2L_2 is secondary leakage inductance referred to the primary side; and a^2R_2 is secondary resistance referred to the primary side. Usually, operation of power transformers is described in terms of these leakage inductances (or leakage reactances) and in terms of resistances, inductances, or reactances all referred to one side of the transformer or the other.

Transformation of impedance. **Figure 6** is the same as Fig. 4 except that the ideal transformer with turn ratio of a:1 has been eliminated and the impedance of the load has been changed from Z to a^2Z. This conversion makes the network of Fig. 6 equivalent to that of Fig. 4 at the input terminals, and it suggests a concept that is quite useful in communications.

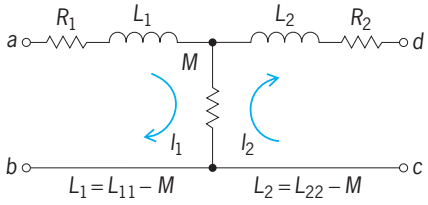

Fig. 5. A conductive network equivalent to the coupled circuits of Fig. 3 ($a = 1$).

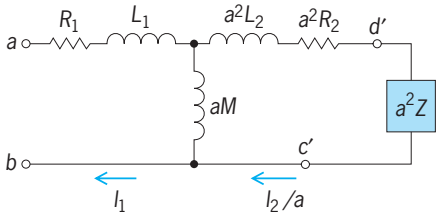

Fig. 6. Transformation of impedance from Fig. 4.

If a load with impedance Z is preceded by an ideal transformer with turn ratio a, the input impedance to the transformer is a^2Z. Thus a load of impedance Z can be made to act like a load with impedance a^2Z by using an ideal transformer with turn ratio a. A useful application is to connect a load of one impedance to an incoming line that has a different terminal impedance through an impedance-matching transformer. By this means a transformer with turn ratio a will match a load with resistance R to a source with resistance a^2R, thereby providing for maximum power transfer to the load.

The preceding paragraph assumes an ideal transformer to provide a perfect impedance match. Figure 6 shows the T network of resistances and inductances that an actual transformer introduces into the network. In practice, the transformer used must be good enough so that these resistances and inductances are acceptable.

Core loss. Neither the equations nor the circuits of this article have taken into account the power loss in a ferromagnetic core. Both eddy-current loss and hysteresis loss may be appreciable; these iron losses are commonly great enough to affect the economics of power transformers, and even to prohibit the use of metal cores with ratio-frequency current. *See* CORE LOSS.

The relations of core loss are not linear and cannot accurately be included in linear equations, but a satisfactory approximation is used with power transformers for which frequency and applied voltage change little if at all. With these restrictions, core loss can be represented by the loss in a resistance shunted around mutual inductance aM of Fig. 4. This approximation of loss is much better than neglecting core loss entirely, and is usual in work with power transformers. *See* CIRCUIT (ELECTRICITY).

Hugh H. Skilling

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Coupling

A mechanical fastening device for connecting the ends of two shafts together. There are three major coupling types: rigid, flexible, and fluid.

Rigid coupling. This connection is used only on shafts that are perfectly aligned. The flanged-face coupling (**Fig. 1***a*) is the simplest of these. For this type of coupling the flanges must be keyed to the shafts. The clamp, or keyless compression, coupling (Fig. 1*b*)has split cylindrical elements which clamp the shaft ends together by direct compression, through bolts, and by the wedge action of conical sections. This coupling is generally used on line shafting to transmit medium or light loads. *See* MACHINE KEY.

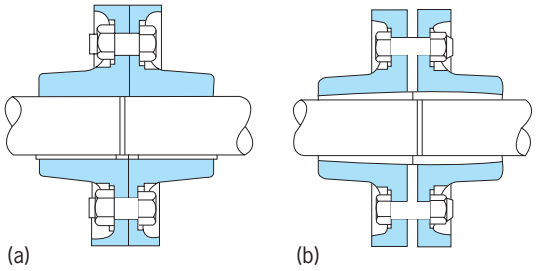

Fig. 1. Rigid coupling. (*a*) Flanged-face coupling. (*b*) Clamp coupling. (*After E. A. Avallone and T. Baumeister III, eds., Marks' Standard Handbook for Mechanical Engineers, 9th ed., McGraw-Hill, 1987*)

Flexible coupling. This connection is used for shafts which are misaligned either laterally or angularly. It also absorbs the impact due to fluctuations in shaft torque or angular speed. The Oldham, or double-slider, coupling (**Fig. 2***a*) may be used to connect shafts that have only lateral misalignment. Because

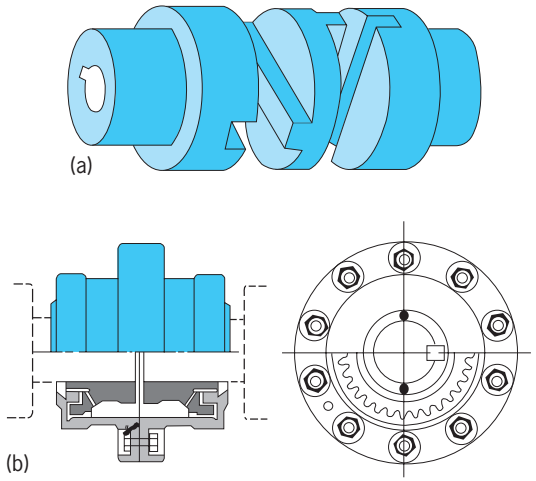

Fig. 2. Flexible coupling. (*a*) Oldham (double-slider) coupling. (*b*) Geared "fast" flexible coupling. (*After E. A. Avallone and T. Baumeister III, eds., Marks' Standard Handbook for Mechanical Engineers, 9th ed., McGraw-Hill, 1987*)

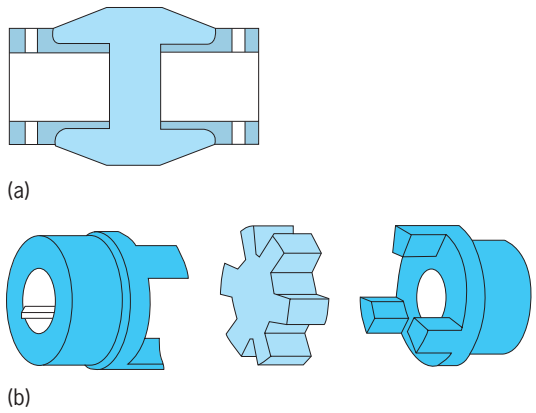

Fig. 3. Rubber flexible coupling. (*a*) Shear type. (*b*) Compression type. (*After E. A. Avallone and T. Baumeister III, eds., Marks' Standard Handbook for Mechanical Engineers, 9th ed., McGraw-Hill, 1987*)

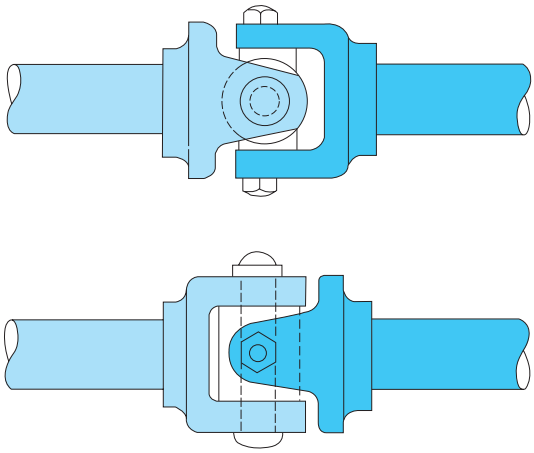

Fig. 4. Hooke's universal joint. (*After E. A. Avallone and T. Baumeister III, eds., Marks' Standard Handbook for Mechanical Engineers, 9th ed., McGraw-Hill, 1987*)

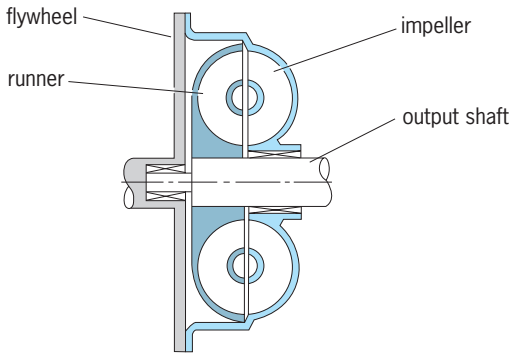

Fig. 5. Fluid coupling. (*After E. A. Avallone and T. Baumeister III, eds., Marks' Standard Handbook for Mechanical Engineers, 9th ed., McGraw-Hill, 1987*)

the tongues move about in the slots, the coupling must be well lubricated and can be used only at low speeds. The geared "fast" flexible coupling (Fig. 2*b*) uses two interior hubs on the shafts with circumferential gear teeth surrounded by a casing having internal gear teeth to mesh and connect the two hubs. Considerable misalignment can be tolerated because the teeth inherently have little interference. This completely enclosed coupling provides means for better lubrication, and is thus applicable for higher speeds. The rubber flexible coupling (**Fig. 3***a*) is used to transmit the torque through a comparatively soft rubber section acting in shear. The type shown in Fig. 3*b* loads the intermediate rubber member in compression. Both types are recommended for light loads only.

Universal joint. This is a flexible coupling for connecting shafts with much larger values of misalignment than can be tolerated by the other types of flexible coupling. Shaft angles up to 30° may be used. The initial universal joint, credited to Robert Hooke (**Fig. 4**), was a swivel arrangement by which two pins at right angles allowed complete angular freedom between two connected shafts. However, it suffers a loss in efficiency with increasing angles. *See* UNIVERSAL JOINT.

Fluid coupling. This type has two basic parts: the input member, or impeller, and the output member, or runner (**Fig. 5**). There is no mechanical connection between the two shafts, power being transmitted by kinetic energy in the operating fluid. The impeller is fastened to the flywheel and turns at engine speed. As this speed increases, fluid within the impeller moves toward the periphery because of centrifugal force. The circular shape of the impeller directs the fluid toward the runner, where its kinetic energy is absorbed as torque delivered by shaft. The positive pressure behind the fluid causes flow to continue toward the hub and back through the impeller. The toroidal space in both the impeller and runner is divided into compartments by a series of flat radial vanes. *See* FLUID COUPLING; SHAFTING. Y. S. Shin

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Cover crops

Unharvested crops grown to improve soil quality or enhance pest management. Increasing numbers of farmers plant cover crops as a means of conserving soil, enhancing production, and reducing off-farm inputs. The sustainable agriculture movement has been a driving force for the increased use of cover crops. Cover cropping can accomplish a wide range of desired benefits, although there can be some drawbacks.

Benefits. Cover crops are often used as a means of preventing soil erosion (**Fig. 1**). Soil erosion is detrimental due to the loss of nutrients and crop rooting depth. Also, the washed-away sediment pollutes streams, rivers, and estuaries and increases the flooding potential downstream.

Cover crops play a vital role in controlling erosion by (1) shielding the soil surface from the impact of falling raindrops; (2) holding soil particles in place; (3) preventing crust formation; (4) improving the soil's capacity to absorb water; (5) slowing the velocity of runoff; and (6) removing subsurface water between storms through transpiration.

Fig. 1. Early spring view of a rolling corn field protected by stalk residues and domestic ryegrass covers.

Cover crops also improve soil structure and water infiltration, both of which are important for crop growth. Production of polysaccharide gums by plant roots and associated microorganisms increases aggregation of soil particles because the gums act as glues to bind particles together. Aggregation of soil particles improves soil tilth, aeration, and drainage of water. Cover crops with a higher carbon-to-nitrogen (C/N) ratio, such as grasses, usually promote greater polysaccharide production and thus greater aggregate stability than do cover crops with lower carbon-to-nitrogen ratios, such as legumes. Cover crops also improve soil structure by adding organic matter. In northern areas it is not unusual for a cover crop like annual ryegrass (*Lolium multiflorum*) to return 4 or more tons (dry weight) of organic matter to the soil. In warmer areas, 5 or more tons can be produced.

Legume cover crops add nitrogen to the soil, which can then be used by crops. They are able to utilize nitrogen from the atmosphere through a symbiotic relationship with bacteria in the genus *Rhizobium*. Through this process, called nitrogen fixation, these bacteria form nodules on legume roots (**Fig. 2**), where gaseous nitrogen is converted into nitrate or ammonium forms usable by plants. Because rhizobia occur in low levels naturally in most soils, the bacteria must be placed on legume seed before planting, either by the seed mill or the farmer; this process is called inoculation. After incorporation of the plants into the soil, the nitrogen becomes available to succeeding crops or to nearby trees or vines. Up to 150–200 lb (68–90 kg) of nitrogen can be made available by sowing legume cover crops. In many cases, legumes can fully satisfy the cash crop's demand for nitrogen.

Fig. 2. Nodules created by nitrogen-fixing rhizobia on berseem clover. Most of the nitrogen fixed in the roots is transported to the tops. (*Photo by Jack Kelly Clark*)

Insect, weed, and nematode management can be affected by cover cropping. Cover crops provide a food source to many beneficial insect species—either nectar, pollen, or other insects. In turn, these beneficial insects may feed on adjacent crop pests. For example, adult parasitic wasps feed on nectar and pollen, while their larvae feed on and kill insect pests, such as aphids. Predator insects, such as lady beetles and green lacewings, are frequently attracted to the cover crop to feed on soft-bodied insects. Many cover crops can effectively suppress weed emergence by shading the weeds or by releasing naturally occurring plant toxins in a process called allelopathy. Species commonly used to suppress weed growth include annual ryegrass, cereal rye (*Secale cereale*), and vetches (*Vicia* spp.). Nematode populations can also be reduced by the use of cover crops. For example, some varieties of cowpea (*Vigna unguiculata*) have been shown to suppress root-knot nematode (*Meloidogyne incognita*).

Potential disadvantages. Cover crops require knowledge and management to attain the desired benefit. If they are not properly selected or managed, there are drawbacks to their use, including depletion of soil moisture, competition with the adjacent crop (when present) for soil moisture and nutrients, increased frost hazard in orchards and vineyards, increased insect, nematode, and weed pests, and added costs to purchase and plant seeds.

Uses and management. Cover crops can be readily grown in humid climates and in arid and semiarid climates where irrigation provides sufficient water that cover crops do not rob the soil of needed moisture. Cover crops that reseed themselves are often sown on rangeland. Long-lived grasses such as blue grama (*Bouteloua gracilis*), buffalo grass (*Buchloë dactyloides*), and the blue-stems (*Andropogon* sp.) have been widely seeded on many different soil types in the Great Plains. In California, annual clovers (*Trifolium* spp.), annual medics (*Medicago* spp.), soft chess (*Bromus hordeaceus*), annual fescue (*Vulpia myuros*), and several perennial grass species are used on rangelands.

In cold climates, many common spring-sown crops can be advantageously used as cover crops. Spring oats (*Avena sativa*) sown in the fall are killed by cold weather but provide excellent ground cover during the fall and winter. Buckwheat (*Fagopyrum esculentum*) can be used as a cover crop to protect the soil and as a smother crop to control weeds; it also improves the soil upon incorporation. Cereal rye is widely seeded as a cover crop following the harvest of the main crop. In mild climates, summer crops such as sorghum-sudangrass (*Sorghum vulgare* × *S. sudanense*) are sometimes planted in late summer, then are killed by frost in late autumn. This practice

Fig. 3. California native perennial grass mixture planted in a drip-irrigated San Joaquin Valley vineyard. This cover crop goes dormant in the summer, reducing competition with the vines for water and nutrients. (*Photo by Jack Kelly Clark*)

protects the soil, suppresses weeds, and scavenges excess soil nitrogen to prevent it from leaching below the root zone.

Cover cropping in orchard and vineyard middles is often easier than in annual crops because they are grown between the crop spatially rather than temporally. A wide range of species, mixes, and management systems are used in orchards and vineyards. For maximum soil protection and wheel traction, perennial grasses such as tall fescue (*Festuca arundinacea*) or ryegrass (*Lolium perenne*) can be used, although these species use substantial soil moisture. In the arid west, some tree and vine crop growers sow native perennial grass species such as California brome (*Bromus carinatus*) and blue wildrye (*Elymus glaucus*) [**Fig. 3**]. These species offer excellent soil protection and go dormant in the summer, using little water. Reseeding annual clovers or grasses are also sown in no-till orchards and vineyards. Many organic growers sow high-biomass legume species such as vetches, clovers, or field peas (*Pisum sativa*), sometimes in combination with cereal cover crops. Upon incorporation, these "green manure" cover crops add nitrogen to the soil for use by the trees or vines. *See* AGRICULTURAL SOIL AND CROP PRACTICES; NITROGEN FIXATION; SOIL CONSERVATION.

Chuck Ingels; Paul J. Zwerman

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Cowpea

The legume *Vigna unguiculata* ssp. *unguiculata*, also called southern pea, blackeye pea, or blackeye bean (United States), or niébé (French-speaking Africa). It is an important source of dietary protein for human consumption and of animal feed in the tropics, especially in Africa, Brazil, and India where cowpeas are grown mostly as a subsistence crop for home consumption and are not sold in markets. The cowpea is adapted to hotter, more arid climates and more infertile soils than other food legume crops. Its symbiotic nitrogen-fixing abilities help maintain soil fertility in peasant cropping systems. Over 65% of the cowpea crop [2.5 million metric tons (5.5 billion pounds) from 8 million hectares (19.8 million acres)] is produced in Africa, Nigeria and Niger producing 50% of the world supply. Brazil produces over 25%. The United States is the only developed country producing large amounts of cowpea. Yields in developing countries are only about 551–882 lb/ha (250–400 kg/ha), principally because of insect infestations but also because of diseases and low soil fertility. In the United States, average yields reach 4410 lb/ha (2000 kg/ha) because of heavier use of inputs, and are comparable to yields of other food legumes.

Origin. The cowpea was domesticated in Africa from one of several wild taxa that belong to the same species as the cultivars and are classified as *V. unguiculata* ssp. *dekindtiana*. The wild relatives are widely distributed in the Sahel, eastern and southern Africa. The precise location of domestication is not known, although northern Nigeria and Benin, Ethiopia, and Botswana have been suggested as possible areas. At least by 2000 years ago, the cowpea was introduced into India, probably with African cereals such as pearl millet and sorghum. In Asia, new cultivar types were selected in addition to the original unguiculata type, grown mostly for dry seeds. The sesquipedalis type, also known as the yard-long or asparagus bean, is grown in India, China, and southeastern Asia for its succulent young pods and leaves. Its pods are up to 18–30 in. (45–75 cm) long. The biflora or catjang type is grown for fodder and dry seeds mostly in India and Sri Lanka. An additional, minor cultivar type is the textilis type, which was grown in northern Nigeria but is now rarely found in farm fields. It is a primitive type, grown for its fibers in the inflorescence peduncles. Cowpea was introduced into the Americas (principally Brazil and the United States) through the colonial and slave trades. In the United States, more than 65% of the production originates in California and Texas (dry seeds) and Georgia (fresh peas).

Characteristics. Cultivars generally have dark-green, glabrous, shiny leaves. Growth habits include climbing, prostrate, and bush types. Flowers and pods are distributed mostly above the leaf canopy, in the upper part of the canopy, or throughout the canopy. Flowers are white or violet with various color patterns. Most cultivars are self-pollinating; however, some wild cowpeas in Africa are cross-pollinated, producing larger, odoriferous flowers. Pods can be erect or pendant (hanging from the peduncle). Their shape ranges from straight to curved to coiled (similar to alfalfa pods). Seeds can be kidney or egg shaped, spherical or rhomboid. The seed coat ranges from smooth to wrinkled. Pigmentation covers the entire seed (self-colored) or surrounds the hilum (eye). Colors are mottled, speckled, or solid

and include white, cream, brown, red, pink, green, and black.

Cultivation. In Africa, the savannas of South America, and southern and southeastern Asia, cowpeas are intercropped with cereals (sorghum, pearl millet, maize) or cotton, or in some cases they are grown as a monoculture. In very dry areas, such as northeastern Brazil, they are grown with spineless forage cactus (*Opuntia* spp.). In the moist Amazon areas of Peru and Brazil, climbing cowpeas are grown. In rubber and tea plantations in Asia, cowpeas are grown as a cover crop or a green manure. In the United States, they are grown in monocultures. Breeders are developing cultivars with improved growth habits; greater resistance to pests, diseases, and parasitic plants (*Striga* and *Alectra*); better tolerance to heat and drought stress; and enhanced symbiotic nitrogen fixation. Novel, low-cost methods are being developed to control cowpea seed weevils in developing countries that do not have adequate pesticide use and crop management programs. *See* LEGUME FORAGES; MULTIPLE CROPPING.

Utilization and nutritional qualities. The most important utilization of cowpeas is that of seeds, whether mature (dry seeds or blackeye bean) or immature (southern pea). Seeds are marketed as a dry pack or as a canned or frozen product. Other uses, such as for the yard-long beans, are becoming more available, especially in specialty food markets. Cowpea seeds contain 23–30% protein. As with other food legumes, these proteins are deficient in the essential amino acids methionine and cystine, but they complement cereal proteins, which are deficient in lysine. *See* LEGUME. P. Gepts

Diseases. Diseases caused by fungi and viruses are primary constraints in cowpea production. Roots can be destroyed by the fungi *Rhizoctonia solani*, *Sclerotium rolfsii*, and *Pythium* sp. *Cercospora cruenta* and *C. canescens* cause leaf spots, which sometimes result in defoliation of mature cowpea plants. Stem anthracnose and pod spot are caused by *Colletotrichum* sp. *Fusarium oxysporum* infects vascular tissue and causes wilting.

Virus diseases of cowpea are characterized by leaf discolorations (mottle, mosaic, chlorosis), stunted growth, and reduced yields. Malformed leaves and distorted growth are caused by cowpea mosaic and southern bean mosaic viruses. Although cucumber mosaic virus causes a very mild disease in cowpea, it frequently occurs in a mixed infection with blackeye cowpea mosaic virus and results in a synergistic interaction with severely stunted plants (see **illus.**). *See* PLANT VIRUSES AND VIROIDS.

The bacterium *Xanthomonas vignicola* can cause both leaf blight and stem canker. In sandy soils a root-knot disease caused by nematodes (*Meloidogyne* sp.) can be particularly injurious to general plant growth.

The most common method to control cowpea diseases is biological. Resistant cultivars are known for many of the diseases, and chemical control usually is not applied. Breeding programs have attempted to incorporate resistance to the major diseases into single cowpea lines. *See* BREEDING (PLANT); PLANT PATHOLOGY. Cedric W. Kuhn

Severe mosaic and stunting of cowpea caused by a mixed infection of cucumber mosaic virus and blackeye cowpea mosaic virus. (*From G. Pio-Ribeiro, S. D. Wyatt, and C. W. Kuhn, Cowpea stunt: A disease caused by synergistic interaction of two viruses, Phytopathology, 68:1260–1265, 1978*)

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Coxsackievirus

A large subgroup of the genus *Enterovirus* in the family Picornaviridae. The coxsackieviruses produce various human illnesses, including aseptic meningitis, herpangina, pleurodynia, and encephalomyocarditis of newborn infants. *See* ENTEROVIRUS; PICORNAVIRIDAE.

Coxsackieviruses measure about 28 nanometers in diameter; they resemble other enteroviruses in many biological properties, but differ in their high pathogenicity for newborn mice. Inapparent infections can be induced in adult mice, chimpanzees, and cynomolgus monkeys. Some types, especially A7, produce severe central nervous system lesions in monkeys. Coxsackieviruses of group A and group B differ in their effects on suckling mice; those of group A produce widespread myositis in skeletal muscle, whereas group B viruses produce necrosis of embryonic fat pads (brown fat). At least 23 antigenically distinct types in group A are now recognized, and 6 in group B. Several types (B1–6, A9) grow well in tissue culture. *See* TISSUE CULTURE.

After incubation for 2–9 days, during which the virus multiplies in the enteric tract, clinical manifestations appear which vary widely.

Diagnosis is by isolation of virus in tissue culture or infant mice. Stools are the richest source of virus. Neutralizing and complement-fixing antibodies form during convalescence and are also useful in diagnosis. *See* ANTIBODY; COMPLEMENT-FIXATION TEST; NEUTRALIZING ANTIBODY.

The coxsackieviruses have worldwide distribution. Infections occur chiefly during summer and early fall, often in epidemic proportions. Spread of virus, like that of other enteroviruses, is associated with family contact and contacts among young children. *See* ANIMAL VIRUS; VIRUS CLASSIFICATION.

Joseph L. Melnick; M. E. Reichmann

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CPT theorem

A fundamental ingredient in quantum field theories, which dictates that all interactions in nature, all the force laws, are unchanged (invariant) on being subjected to the combined operations of particle-antiparticle interchange (so-called charge conjugation, *C*), reflection of the coordinate system through the origin (parity, *P*), and reversal of time, *T*. The operations may be performed in any order; *TCP*, *TPC*, and so forth, are entirely equivalent. If an interaction is not invariant under any one of the operations, its effect must be compensated by the other two, either singly or combined, in order to satisfy the requirements of the theorem. *See* NONRELATIVISTIC QUANTUM THEORY; QUANTUM FIELD THEORY.

The *CPT* theorem appears implicitly in work by J. Schwinger in 1951 to prove the connection between spin and statistics. Subsequently, G. Lüders and W. Pauli derived more explicit proofs, and it is sometimes known as the Lüders-Pauli theorem. The proof is based on little more than the validity of special relativity and local interactions of the fields. The theorem is intrinsic in the structure of all the successful field theories. *See* QUANTUM STATISTICS; RELATIVITY; SPIN (QUANTUM MECHANICS).

Significance. *CPT* assumed paramount importance in 1957, with the discovery that the weak interactions were not invariant under the parity operation. Almost immediately afterward, it was found that the failure of *P* was attended by a compensating failure of *C* invariance. Initially, it appeared that *CP* invariance was preserved and, with the application of the *CPT* theorem, invariance under time reversal. Then, in 1964 an unmistakable violation of *CP* was discovered in the system of neutral *K* mesons. *See* PARITY (QUANTUM MECHANICS).

One question immediately posed by the failure of parity and charge conjugation invariance is why, as one example, the π^+ and π^- mesons, which decay through the weak interactions, have the same lifetime and the same mass. It turns out that the equality of particle-antiparticle masses and lifetimes is a consequence of *CPT* invariance and not *C* invariance alone. A casual proof can be obtained by applying the *CPT* operation to a free particle. This results in an antiparticle with precisely the same momentum and total energy and therefore, from special relativity, the same rest mass as the original particle. This obvious result can be generalized heuristically to include equality of lifetimes by including an imaginary component in the mass.

Experimental tests. The validity of symmetry principles such as the *CPT* theorem should not be assumed beyond the limits of the experimental tests. In that the *CPT* theorem predicts the equality of particle-antiparticle properties such as mass, lifetime, and the magnitude of the magnetic moment, measurements which compare these quantities constitute tests of the theorem. For example, it is known from studies of the neutral *K*-meson system that the mass of the K^0 is the same as the $\overline{K^0}$ to within 1 part in 10^{18}. The magnitude of the magnetic moment of the electron is known to be the same as the positron to 1 part in 10^{10}. The lifetime of the π^+ and π^- mesons is the same to better than 0.1%, and so with the μ^+ and μ^- mesons.

The discovery of *CP* violations in 1964 in the system of neutral *K* mesons reopened the question of *CPT* invariance. That system has been analyzed in great detail, and it is known now that at least 90% of the observed *CP* violation is compensated by a violation of time-reversal invariance. That is, not more than 10% of the observed *CP* violation can be credited to a violation of *CPT*. These are all experimental limits and do not in any way suggest a violation of *CPT*.

The remarkable sensitivity of the *K* mesons to possible departures from *CPT* invariance, as demonstrated by the precision with which the masses of the K^0 and $\overline{K^0}$ are known to be the same, suggests that this system is perhaps the best one for further experimental tests of the validity of the *CPT* theorem. *See* ELEMENTARY PARTICLE; MESON; SYMMETRY LAWS (PHYSICS).

Val L. Fitch

Crab

The name applied to arthropods in sections Anomura and Brachyura of the reptantian suborder of the order Decapoda, class Crustacea. The term crab is also sometimes used for two species of sucking lice (order Anoplura) which prey upon humans. *Phthirus pubis* is the crab louse which inhabits the pubic region.

Section Anomura includes over 1400 species in 12 different families. Commonly called hermit, king, sand, or mole crabs, the anomurans have lobsterlike abdomens which bend beneath the cephalothorax in crablike manner, setting them apart in a somewhat ill-defined group (see **illus.**).

Section Brachyura encompasses the so-called true crabs, with reduced abdomens folded snugly beneath the cephalothorax. More than 4500 species making up 28 families have been grouped into subsections. In addition to subsections Dromiacea, Gymnopleura, and Oxystomata, the Hapalocarcinidea has one family of the same name containing three

Representative species of an anomuran, the hermit crab.

genera of small crabs with elongated abdomens. Young coral gall crab females settle on coral and become enveloped by growing coral. Small openings in the coral, maintained by the crab's water currents, are large enough to permit passage of miniature males (less than 0.08 in. or 2 mm long) and larvae; the female is stuck there for life.

Subsection Brachygnatha is the largest group of brachyurans and includes the most typical crabs. It has 20 families encompassing over 3700 species which occupy a variety of habitats, undergo different types of development, and exhibit contrasting patterns of behavior. Family Cancridae includes popular edible species of the genus *Cancer*. The Portunidae are the swimming crabs, most of which have paddle-shaped modifications of the terminal segments of their fifth pereopods. This structure, advantageous in swimming, is disadvantageous to the blue crab, *Callinectes sapidus*, in that the outline of the new exoskeleton in the thin paddle allows identification of premolt crabs. Such crabs, called "peelers," are kept in cypress wood floats until they molt, then sold as soft-shelled delicacies; the meat of hard-shelled crabs is also delectable. Blue crab fishing and marketing is one of the largest of the shellfish industries on the Atlantic coast of the United States.

Families Potomidae, Pseudothelphusidae, and Trichodactylidae include freshwater species. These crabs are unique in that, like crayfish, the hatching young resemble adults; there is no series of larval stages culminating in metamorphosis. The only other freshwater crabs are four species in the family Grapsidae. These crabs hatch as zoeae which complete their nonfeeding larval development rapidly, using yolk retained from the eggs for nutrition. One Jamaican species, *Metapaulias depressus*, lives in water contained in bromeliad plants. Other grapsid crabs are marine or semiterrestrial forms which invade fresh waters but must undergo postembryonic development in the sea.

The Gecarcinidae, or land crabs, live in humid, tropical regions. *Cardisoma* and *Gecarcinus* may migrate several kilometers from the sea to dig burrows in moist areas. Although the land crabs retain gills, portions of the moist inner surface of the branchial chamber have become highly vascularized to allow respiration in air. A female returns to the sea only to release larvae as they hatch from the fertilized egg mass she carries.

Most Pinnotheridae are tiny, therefore called pea crabs, and live commensally with other organisms. Different species live in the mantle cavities of oysters or other mollusks, in polychaete tubes or burrows, in branchial cavities of ascidians, in cloacas of sea cucumbers, or in the burrows of other crustaceans. After establishing a commensal habitat, many undergo a molt, after which the exoskeleton remains soft.

Other crabs commonly seen at the coast belong to the family Xanthidae (the mud crabs) or the family Ocypodidae. These are amphibious burrowers. *Uca* spp. (Ocypodidae) are called fiddler crabs because males have one enlarged cheliped, the weight of which may be equal that of the rest of the animal. In different species, this claw is used to beat the ground, creating vibrations to threaten, or to wave, an activity which may be a part of courtship. *Ocypode*, a ghost crab, does not have an enlarged cheliped but makes a variety of sounds by stridulation, rubbing the ridged inner surface of the claw against a tubercle on the base of the same cheliped.

Many of the other brachygnathans are marine and inhabit deeper waters. The best known of these are the Majidae, decorator or spider crabs. The Japanese spider crab, *Macrocheira kaempferi*, can have a span up to 11.9 ft (3.6 m) [body width of 18 in. or 45 cm] between the tips of its chelipeds; it is the largest living crab. *See* ARTHROPODA; CRUSTACEA; DECAPODA (CRUSTACEA); MEROSTOMATA.

Marilyn S. Kerr

Crab Nebula

The Crab Nebula is the remnant of a tremendous stellar explosion witnessed by Chinese astronomers in 1054. The explosion, called a supernova, occurred at a distance of about 2000 parsecs from the Earth (1 parsec = 1.9×10^{13} mi = 3.1×10^{13} km = 3.26 light-years). Even so, it was bright enough to be visible in the daytime sky for 3 weeks and in the nighttime sky for 2 years. *See* SUPERNOVA.

Neutron star. The Crab now consists of three components. At the heart of the nebula is what is left of the core of the Crab's giant stellar progenitor. This neutron star has twice the mass of the Sun concentrated into an object only about 20 km (12 mi) across, giving it a density of 10^9 tons per cubic centimeter. The neutron star is spinning at 30 times a second, whipping its powerful (10^8 tesla) magnetic field around with it. Radiation formed in this extreme environment is concentrated into two intense beams directed away from the neutron star's two magnetic poles. As these beams sweep past the direction of the Earth like the beam from a lighthouse, the star appears to wink on and off, earning it the name "pulsar." The Crab pulsar has been seen in all parts of the electromagnetic spectrum from gamma rays through radio waves. *See* CERENKOV

RADIATION; GAMMA-RAY ASTRONOMY; NEUTRON STAR; PULSAR; RADIO ASTRONOMY; X-RAY ASTRONOMY.

Synchrotron nebula. The Crab pulsar's rotational period is slowing by 34 nanoseconds a day, and as it slows it loses 5×10^{31} joules per second, or 100,000 times more power than is radiated away by the Sun. Most of this energy is carried away from the pulsar by a wind of electrons and positrons moving at close to the speed of light. The wind feeds a vast cloud of highly relativistic particles about 3 parsecs × 2 parsecs in size. This cloud is called the Crab synchrotron nebula because, as the particles spiral through the nebula's $3–5 \times 10^{-8}$ tesla magnetic field, they give off the sort of radiation emitted by a synchrotron particle accelerator. The synchrotron nebula is bright in x-rays and ultraviolet radiation as well as the lower-energy visible and radio parts of the spectrum. *See* SYNCHROTRON RADIATION.

Gaseous filaments. The third component of the Crab is a complex of filaments made up of gas ejected by the explosion itself. The filaments, which together have a mass of about twice that of the Sun, are moving away from the site of the explosion at up to around 1500 km/s (900 mi/s). This gas traps the synchrotron "bubble." The pressure of the synchrotron bubble in turn pushes on and compresses the gas into filaments, much like the droplets in an inverted bottle of oil and vinegar salad dressing. Together the mass of the neutron star and the mass of the filaments add up to about four times that of the Sun, yet the understanding of supernova explosions is that the star which exploded must have had at least twice this much mass. The remaining mass from the progenitor may be distributed in a very diffuse cloud of expanding ejecta surrounding the bright nebula (see **illus.**).

The filaments are ionized by ultraviolet radiation from the synchrotron nebula, causing them to glow like the gas in a fluorescent light bulb. Virtually all of the atoms in the universe that are more massive than helium (including the atoms of which humans are made) were formed by nuclear reactions in the interiors of stars and then ejected back into space. Analysis of the light from the Crab filaments shows that they are rich in these more massive elements, the chemical legacy of the fusion furnace that powered the Crab's progenitor until the moment of its cataclysmic death. *See* ASTROPHYSICS, HIGH-ENERGY; FLUORESCENCE; NUCLEOSYNTHESIS. Jeff Hester

Crab Nebula, the 900-year-old remnant of a cataclysmic stellar explosion.

Crabapple

A fruit, represented commercially by such varieties as Martha, Hyslop, and Transcendent, comprising hybrids between *Malus baccata* (Siberian crabapple) and *M. domestica* (cultivated apple).

Trees may grow to 40 ft (12 m) and are very hardy. At one time the fruit of the crabapple was esteemed because of its high pectin content and its usefulness in jam and jelly manufacture. With the introduction of commercial pectin preparations, demand for the crabapple declined sharply. Except for its use as a pickled product, there is little commercial interest in this fruit. *See* PECTIN.

Twenty or more species of Oriental flowering crabapples, some of which produce edible fruit of indifferent quality, are listed. *See* APPLE; FRUIT; FRUIT, TREE; ROSALES. Harold B. Tukey

Cracking

A process used in the petroleum industry to reduce the molecular weight of hydrocarbons by breaking molecular bonds. Cracking is carried out by thermal, catalytic, or hydrocracking methods. Increasing demand for gasoline and other middle distillates relative to demand for heavier fractions makes cracking processes important in balancing the supply of petroleum products.

Thermal cracking depends on a free-radical mechanism to cause scission of hydrocarbon carbon-carbon bonds and a reduction in molecular size, with the formation of olefins, paraffins, and some aromatics. Side reactions such as radical saturation and polymerization are controlled by regulating reaction conditions. In catalytic cracking, carbonium ions are formed on a catalyst surface, where bond scissions, isomerizations, hydrogen exchange, and so on, yield lower olefins, isoparaffins, isoolefins, and aromatics. Hydrocracking, a relative newcomer to the industry, is based on catalytic formation of hydrogen radicals to break carbon-carbon bonds and saturate olefinic bonds. Hydrocracking converts intermediate- and high-boiling distillates to middle distillates, high in paraffins and low in cyclics and olefins. Hydrocracking also causes hydrodealkylation of alkyl-aryl components in heavy reformate to produce benzene and naphthalene. *See* HYDROCRACKING.

Thermal cracking. This is a process in which carbon-to-carbon bonds are severed by the action of heat alone. It consists essentially in the heating of any

fraction of petroleum to a temperature at which substantial thermal decomposition takes place through a thermal free-radical mechanism followed by cooling, condensation, and physical separation of the reaction products. *See* FREE RADICAL.

There are a number of refinery processes based primarily upon the thermal cracking reaction. They differ primarily in the intensity of the thermal conditions and the feedstock handled.

Visbreaking is a mild thermal cracking operation (850–950°F or 454–510°C) where only 20–25% of the residuum feed is converted to mid-distillate and lighter material. It is practiced to reduce the volume of heavy residuum which must be blended with low-grade fuel oils.

Thermal gas-oil or naphtha cracking is a more severe thermal operation (950–1100°F or 510–593°C) where 45% or more of the feed is converted to lower molecular weight. Attempts to crack residua under these conditions would coke the furnace tubes.

Steam cracking is an extremely severe thermal cracking operation (1100–1400°F or 593–760°C) in which steam is used as a diluent to achieve a very low hydrocarbon partial pressure. Primary products desired are olefins such as ethylene, propylene, and butadiene.

Fluid coking is a thermal operation where the residuum is converted fully to gas-oil products boiling lower than 950°F (510°C) and coke. The thermal conversion is carried out on the surface of a fluidized bed of coke particles.

The fluid coking process has been extended to include gasification of the product coke in a vessel that is heat-integrated into the plant. Low-Btu gas which can be desulfurized is produced rather than a product coke stream.

Delayed coking is a thermal cracking operation wherein a residuum is heated and sent to a coke drum, where the liquid has an infinite residence time to convert to lower-molecular-weight hydrocarbons which distill out of the coke drum, and to coke which remains in the drum and must be periodically removed.

In fluid coking and delayed coking, there is total conversion of the very heavy high-boiling end of the residuum feed.

Although there are many variations of visbreaking and thermal cracking, most commonly a feedstock that boils at higher temperatures than gasoline is pumped at inlet pressures of 75–1000 pounds per square inch gage (0.52–6.9 megapascals) through steel tubes so placed in a furnace as to allow gradual heating of the coil to temperatures in the range 850–1100°F (454–593°C). The flow rate is controlled to provide sufficient time for the required cracking to lighter products; the time may be extended by subsequently passing the hot products through a reaction chamber that is maintained at a high temperature. To achieve optimum process efficiency, part of the overhead product ordinarily is returned to the cracking unit for further cracking (**Fig. 1**).

Crude oils differ in their compositions, both in molecular weight and molecular type of hydrocarbon. Since refiners must make products in harmony with market demand, they often need to alter the molecular structure of the hydrocarbons. The cracking of heavy distillates and residual oils increases the yields of gasoline and the light intermediate distillates used primarily as domestic heating oils, as well as providing low-molecular-weight olefins needed for the manufacture of chemicals and polymers.

Beginning in 1912, thermal cracking proved for many years to be eminently suitable for this purpose. During the period 1920–1940, more efficient automobile engines of higher compression ratios were developed. These engines required higher-octane-number gasolines, and thermal cracking operations in the United States were expanded to meet this need. Advantageously, thermal cracking reactions produce olefins and aromatics, leading to gasolines generally of higher octane number than those obtained by simple distillation of the same crude oils. The general nature of the hydrocarbon products and the basic mechanism of thermal cracking is well described by the free-radical theory of the pyrolysis of hydrocarbons. *See* PYROLYSIS.

In the early 1930s, the petrochemical industry began its growth. Olefinic gases from thermal cracking operations, especially propylene and butylenes, were used as the chief raw materials for the production of aliphatic chemicals. Stimultaneously, practical catalytic processes were invented for polymerizing propylene and butylenes to gasoline components, and for dimerizing and hydrogenating isobutylene to isooctane (2,2,4-trimethylpentane), the prototype 100-octane fuel. Just prior to World War II, the alkylation of light olefins with isoparaffins to produce unusually high-octane gasoline components was discovered and extensively applied for military aviation use. These developments resulted in intense engineering efforts to bring thermal cracking to maximum efficiency, as exemplified by a number of commercial processes made available to the industry.

Since World War II, thermal cracking has been largely supplanted by catalytic cracking, both for the manufacture of high-octane gasoline and as a source of light olefins. It is, however, still used widely for the mild cracking of heavy residues to reduce their viscosities and for the final cracking of low-grade residuum.

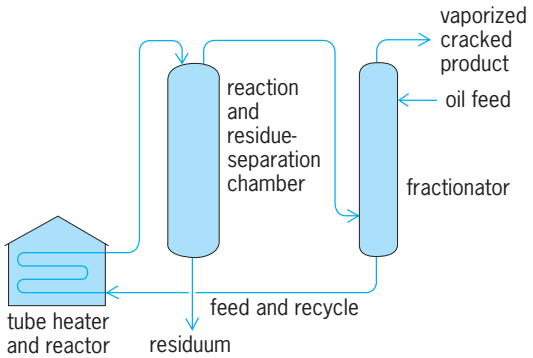

Fig. 1. Thermal cracking unit.

Since carbon-to-hydrogen bonds are also severed in the course of thermal cracking, two hydrogen atoms can be removed from adjacent carbon atoms in a saturated hydrocarbon, producing molecular hydrogen and an olefin. This reaction prevails in ethane and propane cracking to yield ethylene and propylene, respectively. Methane cracking is a unique case wherein molecular hydrogen is obtained as a primary product and carbon as a coproduct. These processes generally operate at low pressures and high temperatures and in some cases utilize regenerative heating chambers lined with firebrick, or equipment through which preheated refractory pebbles continuously flow. Such conditions also favor the production of aromatics and diolefins from normally liquid feedstocks and are applied commercially to a limited extent despite relatively low yields of the desired products.

Catalytic cracking. This is the major process used throughout most of the world oil industry for the production of high-octane quality gasoline by the conversion of intermediate- and high-boiling petroleum distillates to lower-molecular-weight products. Many aspects involve proprietary technology. In general, oil heated to within the lower range of thermal cracking temperatures (850–1025°F or 454–551°C) reacts in the presence of an acidic inorganic catalyst, typically a silica-alumina zeolite, under low pressures (10–35 psig or 70–240 kilopascals). Gasoline of much higher octane number is obtained than from thermal cracking, a principal reason for the widespread adoption of catalytic cracking. All nonvolatile carbonaceous materials are deposited on the catalyst as coke and are burned off during catalyst regeneration. *See* ZEOLITE.

In contrast to thermal cracking, low-grade residual oils are not generally processed, because excessive amounts of coke are deposited on the catalyst, and inorganic components of these oils contaminate the catalyst. Preferential feedstocks are heavier-boiling distillate gas oils. However, in recent years with catalyst improvement, residual stocks can also be processed. These stocks are generally good-quality residuals in low concentration as they yield 2–3 barrels of gas oil feed when processed in a full unit.

Many processes have become available to improve catalytic cracking feedstocks. These include hydrogenation, deasphalting (including the Rose process), deep-vacuum distillation, and separations or low-conversion processes such as the Art process. These processes generally reject the poorer components of the feed, such as Conradson carbon (a residual carbon), metallic contaminants, and possibly heavy aromatics. Hydrogenation can upgrade feedstocks by saturation or aromatics, and remove metals to considerably improve catalytic cracker product selectivity.

Catalytic cracking, as conceived by E. J. Houdry in France, reached commercial status in 1936 after extensive engineering development by American oil companies. In its first form, the process used a series of fixed beds of catalyst in large steel cases. Each of these alternated between oil cracking and catalyst-coke burning at intervals of about 10 min and provided for heat and temperature control.

Successful operation led to major engineering improvements, and the goals of much improved efficiency, enlarged capacity, and ease of operation were achieved by two different systems. One employs a moving bed of small pellets or beads of catalyst traveling continuously through the oil-cracking vessel and subsequently through a regeneration kiln. The beads are lifted mechanically or by air to the top of the structure to flow down through the vessels again. This process has two commercially engineered embodiments, the Thermofor (or TCC) and the Houdriflow processes, which are similar in general process arrangements.

These moving-bed processes are limited in size and became technically obsolete. They generally have been replaced by another type of unit, the fluid-solids, as dictated by economic considerations, and no new moving-bed units are being constructed. In the fluid-solids unit, a finely divided powdered catalyst is transported between oil-cracking and air-regeneration vessels in a fluidized state by gaseous streams in a continuous cycle. This system employs the principle of balanced hydrostatic heads of fluidized catalyst between the two vessels. Catalyst is moved by injecting heated oil vapors into the transport line from the regenerator to the reactor, and by injecting air into the transport line from the reactor stripper to the regenerator. Large amounts of catalyst can be moved rapidly; cracking units of total oil intake as great as 180,000 bbl/day (28,800 m^3/day) are in operation (**Fig. 2**).

In both the moving-bed and fluidized systems, the circulating catalyst provides the cracking heat. Coke deposited on the catalyst during cracking is burned at controlled air rates during regeneration; heat of combustion is converted largely to sensible heat of the catalyst, which supplies the endothermic heat of cracking in the reaction vessel.

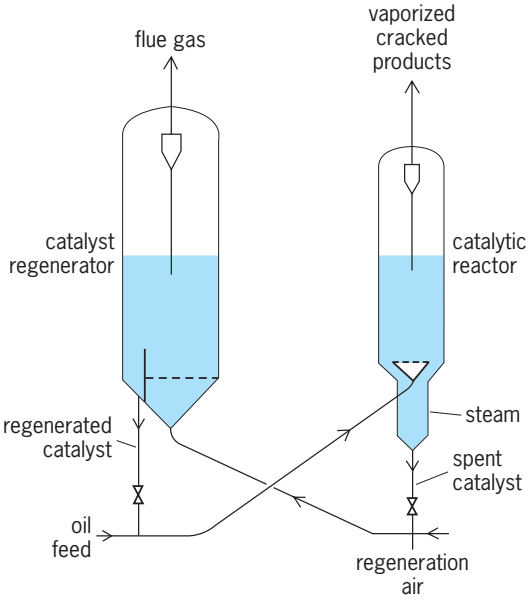

Fig. 2. Fluid catalytic cracking unit.

Gasoline of 90–95 research octane number without tetraethyllead is rather uniformly produced by catalytic cracking of fractions from a wide variety of crude oils, compared with 65–80 research octane number via thermal cracking, the latter figures varying with crude oil source. *See* OCTANE NUMBER.

Although the primary objective of catalytic crackling is the production of maximum yields of gasoline concordant with efficient operation of the process, large amounts of normally gaseous hydrocarbons are produced at the same time. The gaseous hydrocarbons include propylene and butylenes, which are in great demand for chemical manufacture. Isobutane and isopentane are also produced in large quantities and are valuable for the alkylation of olefins, as well as for directly blending into gasoline as high-octane components.

The other chief product is the material boiling above gasoline, designated as catalytically cracked gas oil. It contains hydrocarbons relatively resistant to further cracking, particularly polycyclic aromatics. The ligher portion may be used directly or blended with straight-run and thermally cracked distillates of the same boiling range for use as a diesel oil component and heating oils. Part of the heavier portion can be recycled with fresh feedstock to obtain additional conversion to lighter products. The remainder is withdrawn for blending with residual oils to reduce the viscosity of heavy fuel, or else subjected to a final step of thermal cracking.

Thus, the catalytic cracking process is used in refineries to shift the production of products to match swings in market demand. It can process a wide variety of feeds to different product compositions. For example, light gases, gasoline, or diesel oil can be emphasized by varying process conditions, feedstocks, and boiling range of products (see **table**).

To account for the difference between the product compositions obtained by catalytic and thermal cracking, the mechanism of cracking over acidic catalysts has been investigated intensively. In thermal cracking, free radicals are reaction intermediates, and the products are determined by their specific decomposition patterns. In contrast, catalytic cracking takes place through ionic intermediates, designated as carbonium ions (positively charged free radicals) generated at the catalyst surface. Although there is a certain parallelism between the modes of cracking of free radicals and carbonium ions, the latter undergo rapid intramolecular rearrangement reactions prior to cracking. This leads to more highly branched hydrocarbon structures than those from thermal cracking, and to important differences in the molecular weight distribution of the cracked products. Furthermore, the cracked products undergo much more extensive secondary reactions with the catalyst. *See* REACTIVE INTERMEDIATES.

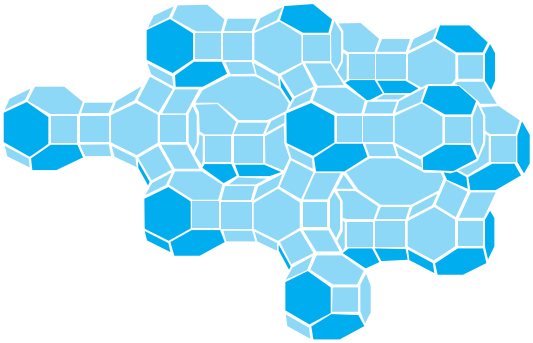

Fig. 3. Model of zeolite type Y.

The catalytic cracking mechanism also favors the production of aromatics in the gasoline boiling range; these reach quite high concentrations in the higher-boiling portion. This characteristic, together with the copious production of branched aliphatic hydrocarbons especially in the lower-boiling portion, is largely responsible for the high octane rating of catalytically cracked gasoline.

Cracking catalysts must have two essential properties: (1) a chemical composition capable of maintaining a high degree of acidity, preferably as readily available hydrogen ions (protons); and (2) a physical structure of high porosity so that these active sites are available for cracking. Mechanical durability is also necessary for industrial use.

Cracking catalysts are essentially silica-alumina compositions. A dramatic improvement in catalytic unit performance occurred with the switch from acid-treated clays (montmorillonite or kaolinite) to synthetic silica-aluminas. After 1960, a new group of aluminosilicates, molecular sieve zeolites, were introduced into the catalyst formulation. These crystalline materials (**Fig. 3**) have cracking activity 50 to 100 times the previous amorphous catalyst. They permit cracking to greater conversion levels, producing more gasoline, less coke, and less gas. *See* ZEOLITE.

As the catalyst particles pass through the reactor regenerator system every 3 to 15 min they are gradually deactivated through loss of active surface area

Representative yield structures for three different processing objectives in catalytic cracking

Process variables	Light gases	Gasoline	Light cycle oil
Feed	Light gas oil	Gas oil	Gas oil
Reactor temperature, °F (°C)	990 (532)	950–990 (510–532)	890–900 (477–482)
Light gases, wt %	4.5	2.8	1.6
Propane/propylene, vol %	15	10.0	7.5
Butane/butylene, vol %	22	16.4	11.2
Gasoline, vol %	46	69.5	32.6
Catalytic diesel oil, vol %	18	10	43.6
Bottom, vol %	5	5	5

by the effect of heat and steam and through contamination by the trace metallic components on the feedstocks, mainly sodium and vanadium. The catalyst particles also undergo mechanical attrition, and fines are lost in the reactor and regenerator gases. To compensate, fresh catalyst is added.

Cracking catalysts have evolved considerably since the mid-1960s, when zeolite crystals were incorporated into the matrix. The sodium and hydrogen ions in the crystals have been replaced with rare-earth elements to provide thermal stability. Surface area of the catalysts has been reduced to provide increased resistance to contaminant metals (Ni and V) and to reduce coke-make of the reaction.

With increased levels of metals on the catalysts, processes have been developed to passivate the effect of the metals. One process involves addition of antimony to the catalytic feedstock. Antimony thus incorporated into the catalyst can reduce hydrogen and carbon yield attributable to the metals.

The new zeolite catalysts have resulted in considerable change in the process itself. The catalysts are made more resistant to thermal degradation primarily by the exchange of sodium and hydrogen ions in the zeolite with rare-earth elements. Regenerator temperatures can be safely raised to the 1350°F (732°C) or greater level. The carbon on regenerated catalyst is reduced to the 0.05 wt % level resulting in improved gasoline yields. In addition, all the carbon monoxide produced at lower generation temperatures (10% concentration in the regenerator flue gas) can be combusted in the regenerator if desired, making for a more efficient recovery of combustion heat and reduced atmospheric pollution. The effluent CO concentration can be reduced to less than 0.05 vol %, if desired.

Additives have been developed to improve the efficiency of regeneration. Combustion promoters can ensure complete combustion of the flue gas to CO_2 even at relatively low temperatures of 1250°F (695°C). These promoters generally contain an element from the platinum group in extremely low concentrations, the ppm range. They may be added as separate powders or in admixtures with the catalyst. Other additives are being produced which are claimed to reduce sulfur oxides in the flue gas.

The high-activity zeolite catalyst has permitted units to be designed with all riser cracking (**Fig. 4**), wherein all the cracking reaction takes place in a relatively dilute (less than 2 lb/ft^3 or 3.2 kg/m^3) catalyst suspension in a 2- or 5-s residence time. No dense-bed (10–15 lb/ft^3 or 16–24 kg/m^3) cracking exists in these units.

Many old units are being converted to riser crackers, and virtually all new units feature riser cracking. Some riser crackers are also provided with small dense beds to achieve an optimum yield pattern. Some have riser-type regenerators directed to improve regenerator efficiency.

Over the years, there have been continued improvements to the process in many areas. Feeds available for processing have become much heavier, and include stocks such as vacuum residua containing contaminant metal up to 30–40 ppm. Many catalyst formulations have been developed to increase catalytic activity and to resist the high level of contaminant metals on the catalyst. Levels as high as 10,000 ppm of nickel and vanadium on catalyst have been run commercially. Nickel, vanadium, and copper catalyze detrimental competing dehydrogenation reactions that produce coke and light gases (hydrogen and methane). These heavier stocks have produced larger amounts of coke which must be burned in the regenerator, causing regenerator temperatures to rise. Staged regeneration is reported to permit temperatures as high as 1500°F (815°C), since water of combustion is liberated in the lower-temperature first stage, and thus hydrothermal deactivation is not as much a factor in the second stage. Additionally, facilities have been provided internally and externally to the regenerator to remove excess heat of combustion.

With more emphasis on environmental protection, complex facilities are provided to remove pollutants from the effluent regeneration gases. Third-stage cyclone collectors, electrostatic precipitators, and scrubbers are used commercially to meet regulations.

To conserve energy in the process, flue gas expanders are used in the flue gas system to provide more than enough energy to compress the regeneration air.

With the continuing emphasis on conservation of petroleum resources, catalytic cracking assumed a more important role. The heavier products from a refinery could be replaced by coal-, shale-, or tar-sand-derived products. This creates a need for greater conversion of petroleum to gasoline, with

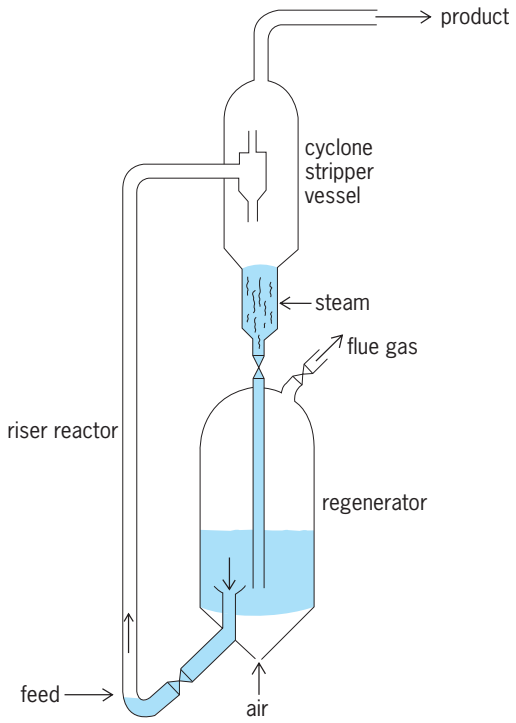

Fig. 4. Riser catalytic cracker in "pure" transfer-line configuration.

emphasis on catalytic cracking since it is the cheapest major conversion process. *See* ALKYLATION (PETROLEUM); AROMATIZATION; COAL LIQUEFACTION; FIRED HEATER; HETEROGENEOUS CATALYSIS; ISOMERIZATION; PETROLEUM PROCESSING AND REFINING. Edward Luckenbach

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Cranberry

The large-fruited American cranberry, *Vaccinium macrocarpon*, a member of the heath family, Ericaceae, is a native plant of open, acid peat bogs in northeastern North America. Selections from the wild have been cultivated since the early nineteenth century. It is an evergreen perennial vine producing runners and upright branches with conspicuous terminal flower buds (**Fig. 1**). *See* ERICALES.

Cultivation. A cranberry bog is made by removing the vegetation on a maple, cedar, or brown-brush swamp, draining it by cutting "shore," "lateral," and "main" ditches, spreading 2 or 3 in. (5 or 8 cm) of sand over the leveled peat, and inserting cuttings of the selected variety of vines through the sand into the peat. Rooting occurs readily within a month, but 3 or 4 years of growth and care are required before the first commercial crop is produced.

Care includes frost protection in spring and fall, now largely provided by solid-set, low-gallonage sprinkler systems which effect protection much quicker and with only a tenth of the water formerly required with flood frost protection. Flooding is used in winter to protect the cranberry vines, not from cold but from desiccation when the root zone is frozen. Early-season floods are used for drowning pest insects.

Because well-tended cranberry bogs continue to produce annual crops for a century or more without replanting, a half-inch (1.3 cm) layer of sand is spread every 3 or 4 years over the vines (by spreading on winter ice when thick enough) to cover the accumulating cranberry leaves on the soil surface where the cranberry girdler insect would otherwise breed and multiply.

Fig. 1. Cranberry uprights in full bloom.

Harvesting. Commercial cranberry growing is confined to Massachusetts, New Jersey, Wisconsin, Washington, and Oregon and to several provinces in Canada. Only about 20% of the cranberries are sold as fresh fruit, and most of these are dry-harvested by machines. Most of the processing fruit is harvested in flood waters, either by machines which pick and deliver the fruit into towed plastic boats or by water reels which detach the berries to float and be driven by wind to shore where they are elevated into bulk trucks. Berries for juice manufacture are usually vine-ripened and deep-frozen for a month or more prior to thawing and extraction. The older form of hand harvest with wood-toothed scoops is now gone, the scoops sometimes being used for ditch-edge picking.

Good-quality fresh cranberries can be stored for several months, refrigerated, with very little loss to decay. Good-quality cranberries can be kept in deep-freeze storage for several years with only minor moisture loss. Frozen berries on thawing are soft and juicy, unlike the firm fresh berry, and must be utilized promptly.

Cranberry bogs in full bloom in early July are a sight to remember because they have 40,000,000 to 50,000,000 white or pink flowers per acre (100,000,000 to 125,000,000 per hectare), each of which must be visited by pollinating honeybees or bumblebees to set the berries. The control of frost and winter injury, the control of weeds, insects, and diseases, and the development of more efficient harvest techniques have seen productivity rise. Supplies of the fruit are ample to satisfy both domestic and export demand.

The special requirements of the cranberry plant for low fertility, acid soil, and winter protection make it a poor choice for home garden cultivation. Chester E. Cross

Diseases. Fruit rots are the most economically important diseases of cranberry. Eight species of fungi cause almost all of the loss attributed to these diseases.

Symptoms of all fruit rots are almost identical, and include the development of a discolored spot on the surface of the fruit; the spot enlarges rapidly until the entire berry is affected (**Fig. 2**). Later the infected berry shrivels into a raisinlike mummy.

Several factors influence the amount of fruit rot that occurs. Higher temperatures during the growing season increase fungal activity and the level of disease. High moisture levels also increase disease, because spore release, spore germination, and penetration of the host plant are all dependent on the presence of adequate moisture. Bruising during harvest, sorting, or packing significantly increases the amount of postharvest fruit rot.

Infection before harvest is common in the production areas with warmer climates and higher rainfall. In such areas fungicide sprays, at midbloom and

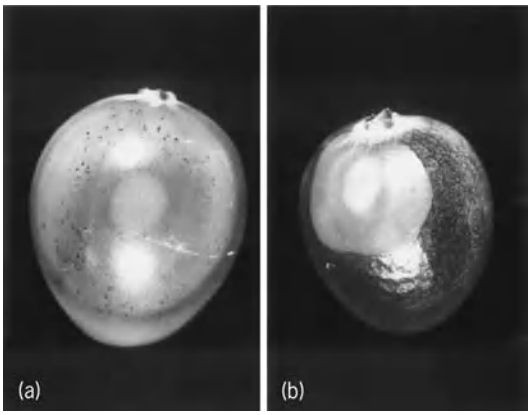

Fig. 2. Symptoms of cranberry disease. (*a*) Early rot symptoms on fruit; black specks in the affected area are the fruiting bodies of the fungus. (*b*) Early symptoms of blotch rot on fruit; all fungal fruit rots appear identical at this stage.

every 14 days until all blossoms have fallen, have provided excellent control.

Practices that reduce fruit rots include rapid drying of the fruit, reducing fruit temperatures to 36–40°F (2–4°C), and reducing the time the fruit remains in storage. Dry harvest results in less disease than wet harvest. *See* PLANT PATHOLOGY. John K. Springer

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Cranial nerve

Any peripheral nerve which has its central nervous system connection with the brain, as opposed to the spinal cord, and reaches the brain through a hole (foramen) in the skull. Nerve fibers are sensory if they carry information from the periphery to the brain, and motor if they carry information from the brain to the periphery. Sensory fibers are classified as somatic sensory if they come from the skin or muscle sense organs, visceral sensory if they come from the viscera, and special sensory if they come from special sense organs such as the eye and ear. Motor fibers are classified as somatic motor if they carry information to somatic striated muscles; general visceral motor if they carry information to glands, smooth muscle, or cardiac muscle; and special visceral motor if they carry information to visceral striated muscle. A cranial nerve may have only one fiber type or several; cranial nerves with several fiber types are called mixed nerves.

In mammals, 12 pairs of cranial nerves, numbered I through XII, are usually described (see **table**). However, mammals also have an anterior unnumbered nerve, the terminal nerve. Many vertebrates also have two pairs of lateral line nerves (unnumbered), and lack discrete nerves XI and XII. *See* BRAIN; NERVOUS SYSTEM (VERTEBRATE). Douglas B. Webster

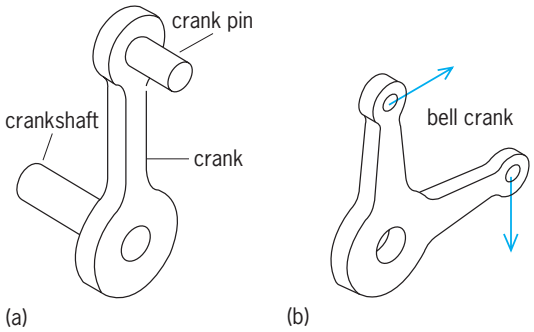
Cranial nerves of vertebrates

Number	Name	Fiber types	Peripheral origin or destination	Vertebrates possessing this nerve
—	Terminal	Somatic sensory	Anterior nasal epithelium	Almost all
I	Olfactory	Special sensory	Olfactory mucosa	All
—	Vomeronasal	Special sensory	Vomeronasal mucosa	Almost all
II	Optic	Special sensory	Retina of eye	All
III	Oculomotor	Somatic motor	Four extrinsic eye muscles	All
IV	Trochlear	Somatic motor	One extrinsic eye muscle	All
V	Trigeminal	Special visceral motor	Muscles of mandibular arch derivative	All
		Somatic sensory	Most of head	All
VI	Abducens	Somatic motor	One extrinsic eye muscle	All
—	Anterior lateral line	Special sensory	Lateral line organs of head	Fish and larval amphibians
VII	Facial	Special visceral motor	Muscles of hyoid arch derivative	All
		General visceral motor	Salivary glands	All
		Somatic sensory	Small part of head	All
		Visceral sensory	Anterior pharynx	All
		Special sensory	Taste, anterior tongue	All
VIII	Vestibulocochlear	Special sensory	Inner ear	All
	Posterior lateral line	Special sensory	Lateral line organs of trunk	Fish and larval amphibians
IX	Glossopharyngeal	Special visceral motor	Muscles of third branchial arch	All
		General visceral motor	Salivary gland	All
		Somatic sensory	Skin near ear	All
		Visceral sensory	Part of pharynx	All
		Special sensory	Taste, posterior tongue	All
X	Vagus	Special visceral motor	Muscles of arches 4±6	All
		General visceral motor	Most viscera of entire trunk	All
		Visceral sensory	Larynx and part of pharynx	All
		Special sensory	Taste, pharynx	All
XI	Spinal accessory	Special visceral motor	Some muscles of arches 4–6	Reptiles, birds, mammals
XII	Hypoglossal	Somatic motor	Muscles of tongue and anterior throat	Reptiles, birds, mammals

Crank

In a mechanical linkage or mechanism, a link that can turn about a center of rotation. The crank's center of rotation is in the pivot, usually the axis of a crankshaft, that connects the crank to an adjacent link. A crank is arranged for complete rotation (360°) about its center; however, it may only oscillate or have intermittent motion. A bell crank is frequently used to change direction of motion in a linkage (see **illus.**). *See* LINKAGE (MECHANISM).

In mechanisms where energy input fluctuates, it is



Cranks (*a*) for changing radius of rotation, and (*b*) for changing direction of translation.

sometimes useful to design a crank stoutly, as a form of flywheel, so that considerable rotational energy is stored in it, tending thereby to make the output smoother inasmuch as the inertia of such a crank absorbs peak input and releases energy more smoothly to output.

Input and output cranks of a four-bar linkage may express in their relative angular placement a relation, $\beta = \beta(\alpha)$, which is excellent in control operations where electric current computers would be vulnerable. *See* FOUR-BAR LINKAGE. Douglas P. Adams

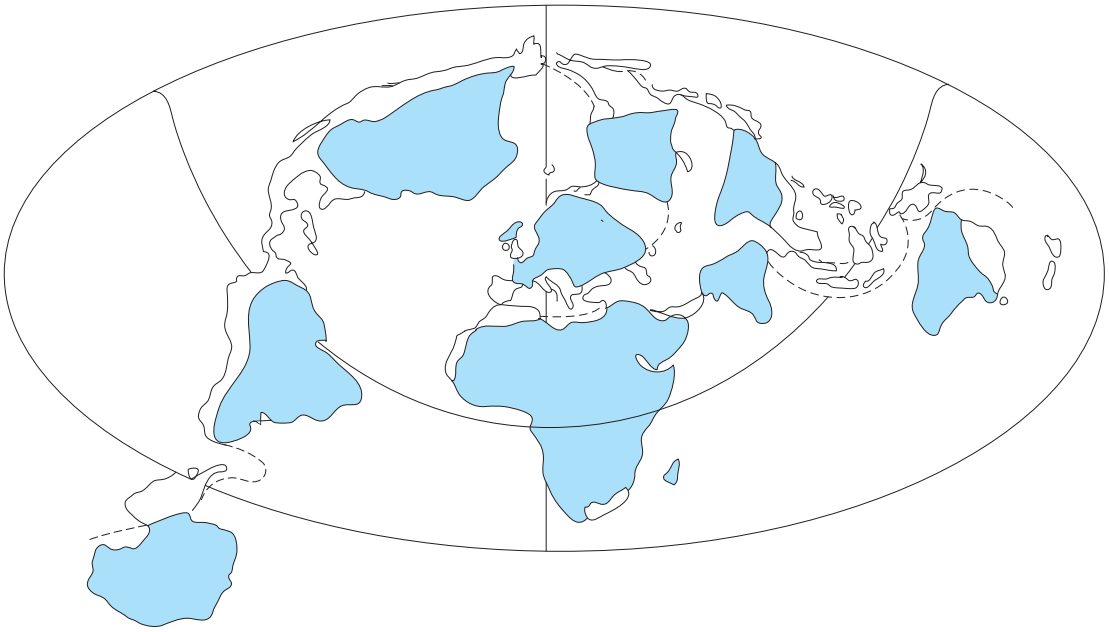
Craton

A large, relatively stable portion of the Earth's crust. Although ocean basins originally were considered low cratons, today the term applies only to continents. Continental (high) cratons are the broad heartlands of continents with subdued topography, encompassing the largest areas of most continents. Cratons experience only broad (epeirogenic) warping and occasional faulting in contrast to the much more structurally mobile or unstable zones of continents, which include mountain ranges, such as the Himalaya, and rift zones, such as those of East Africa. The terminology used today to express these contrasts was first proposed by the German geologist L. Kober in 1921. *Kratogen* referred to stable continental platforms, and *orogen* to mountain or orogenic belts. Later authors shortened the former term to kraton or craton. A complementary term, taphrogen, coined in the 1940s, encompasses the rift structures.

The present North American craton comprises the low continental interior extending from the young Rocky Mountains east to the older Paleozoic Appalachian Mountains, and north to the Paleozoic Franklin mountain belt along the Arctic margin of Canada and Greenland. The Canadian shield is merely that part of the craton where pre-Paleozoic rocks are widely exposed today. In contrast, the remainder of the craton in the plains of western Canada and most of the central United States has little-deformed younger strata at the surface. Such Phanerozoic strata must have covered most of the shield in the past but have been eroded. Most continents have similar large Phanerozoic cratons—for example, all of South America east of the Andes Mountains; Australia west of the Great Dividing Range; Siberia in northern Asia; Europe from the Baltic Sea to the Ural Mountains; all of Africa except the northern and southern edges (orogens) and East Africa (rift valleys); India south of the Himalaya; and all but the Pacific side of Antarctica (see **illus.**). China and Southeast Asia are exceptional today because of the ongoing collision of India with Asia, which has caused intense fragmentation of former cratons there by great, still-active faults. *See* CONTINENTS, EVOLUTION OF; EARTH CRUST.

The recognition of fundamental differences in structural behavior between cratons and orogens long preceded their formal classification. In the 1850s, J. Hall recognized a profound difference between Paleozoic strata in the lowland interior of North America and the Appalachian Mountains. In the mountains, such strata were 10 times thicker, were much deformed, and were metamorphosed in the eastern ranges. Not only were the equivalent strata in the interior still nearly flat-lying and thinner, but there were also subtle contrasts in lithology. Hall postulated a simplistic cause-and-effect relationship, with greater thickness of strata along the eastern continental margin causing subsidence of the crust, which in turn resulted in crumpling of those same strata. In 1873 J. Dana proposed instead that cooling of the earth had caused buckling of the margin between the continental and oceanic crust. The Appalachian region was a downbuckled zone termed a geosyncline, which could accommodate the thick sediments. Eventually this zone could bend downward no more and failed due to compression, which caused the crumpling and elevation of the strata to form the Appalachian Mountains. European geologists also were highlighting the structural contrast between cratons (or forelands) and orogens. Soon the theory of continental drift, proposed in 1910 by F. Taylor and in 1912 by A. Wegener, presented a new mechanism for explaining orogens as the result of the collision between two continental cratons or between a craton and some smaller tectonic element such as an island archipelago. This new explanation eventually evolved into the modern theory of plate tectonics. *See* CONTINENTAL DRIFT; CONTINENTAL MARGIN; OROGENY; PLATE TECTONICS.

Today, cratons are believed to comprise bits of old continental crust, which have had long and complex histories involving the overprinting of many tectonic events. Continents can be viewed as great collages of tectonic elements amalgamated together at different times and in different ways (see illus.). Therefore, the delineation of both cratons and orogens has an important temporal element: What is now a craton may be made up of interlaced orogens of differing ages, each of which may include bits of mangled older cratons within them. For example, 1-billion-year-old orogens may have been fused to or may have cut across orogens of 2-billion-year age, which in turn may surround a cratonic kernel more than 3 billion years old. Returning to the present, most of the North American craton has been a relatively



World map showing Phanerozoic continental cratons (colored areas) of the last 500–600 million years. Each craton is an amalgam of older tectonic elements, which include fragments of complexly interlaced pre-Phanerozoic cratons and orogens. (*After R. Dott and R. L. Batten, Evolution of the Earth, 8th ed., McGraw-Hill, 1988*)

stable element for at least the past 600 million years. Prior to that, the configuration among cratons and orogens was very different, and only the ghosts of some of the fragmentary components of the present continent are discernible. Refined studies of seismic waves that have penetrated the Earth's deep interior indicate that beneath cratons there is anomalous mantle with distinctive compositional or thermal characteristics. Such roots imply a long-term linkage between the evolution of early continental crust and underlying mantle, which somehow contributed to the survival of fragments of cratons older than 2.5–3 billion years. Robert H. Dott, Jr.

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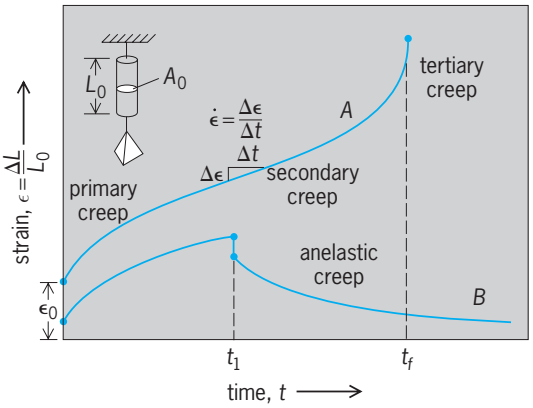
Creep (materials)

The time-dependent strain occurring when solids are subjected to an applied stress. *See* STRESS AND STRAIN.

Kinds of phenomena. Some of the different kinds of creep phenomena that can be exhibited by materials are shown in the **illustration**. The strain $\epsilon = \Delta L/L_0$, in which L_0 is the initial length of a body and ΔL is its increase in length, is plotted against the time t for which it is subjected to an applied stress. The most common kind of creep response is represented by the curve A. Following the loading strain ϵ_0, the creep rate, as indicated by the slope of the curve, is high but decreases as the material deforms during the primary creep stage. At sufficiently large strains, the material creeps at a constant rate. This is called the secondary or steady-state creep stage. Ordinarily this is the most important stage of creep since the time to failure t_f is determined primarily by the secondary creep rate $\dot{\epsilon}_s$. In the case of tension creep, the secondary creep stage is eventually interrupted by the onset of tertiary creep, which is characterized by internal fracturing of the material, creep acceleration, and finally failure. The creep rate is usually very temperature-dependent. At low temperatures or applied stresses the time scale can be thousands of years or longer. At high temperatures the entire creep process can occur in a matter of seconds. Another kind of creep response is shown by curve B. This is the sort of strain-time behavior observed when the applied stress is partially or



Typical creep curves for materials.

completely removed in the course of creep. This results in time-dependent or anelastic strain recovery.

Examples. Creep of materials often limits their use in engineering structures. The centrifugal forces acting on turbine blades cause them to extend by creep. In nuclear reactors the metal tubes that contain the fuel undergo creep in response to the pressures and forces exerted on them. In these examples the occurrence of creep is brought about by the need to operate these systems at the highest possible temperatures. Creep also occurs in ordinary structures. For example, when a bolt is tightened to fasten two parts together, the tension stress in the bolt can be quite high. Over a long period of time the bolt creeps, with the result that the force holding the two parts together diminishes. Another example of stress relaxation caused by creep is found in prestressed concrete beams, which are held in compression by steel rods that extend through them. Creep and stress relaxation in the steel rods eventually leads to a reduction of the compression force acting in the beam, and this can result in failure. *See* PRESTRESSED CONCRETE.

There are many examples of creep that are not related to high technology. The window panes in very old homes or cathedrals are often thicker at the bottom than they are at the top due to creep under the force of gravity. Similarly, when a glacier flows down the side of a mountain the ice undergoes extensive creep deformation.

Mechanism. The mechanism of creep invariably involves the sliding motion of atoms or molecules past each other. In amorphous materials such as glasses, almost any atom or molecule within the material is free to slide past its neighbor in response to a shear stress. In plastics, the long molecular chains can slide past each other only to a limited extent. Such materials typically show large anelastic creep effects (curve B in the illustration).

For crystalline materials, creep deformation also involves the sliding of atoms past each other, but here the sliding can occur only within the cores of crystal dislocations. Thus, creep of metals and ceramics is usually governed by the motion of dislocations. In addition to gliding along their slip planes, the dislocations can climb out of their slip planes by the absorption or emission of lattice vacancies. This process is controlled by the rate at which lattice vacancies diffuse to or from climbing dislocations and gives rise to the strong temperature dependence of creep.

Materials with creep resistance. Understanding of the mechanisms that control creep deformation makes it possible to design materials with superior creep resistance. When solute atoms are added to metals, they are attracted to the strain fields of the dislocations. There they inhibit dislocation motion and in this way improve the creep resistance. Many of the aluminum alloys used for aircraft structures are strengthened in this way. The addition of second-phase particles to alloys is another way to improve the creep resistance. The most effective strengthening phases are oxides, carbides, or intermetallic phases, because they are usually much stronger than the host metal and therefore create strong obstacles to dislocation motion. Materials containing finely dispersed, strong particles of a stable phase are usually very creep-resistant. Nickel-based superalloys, used in gas-turbine engines, derive their creep resistance from these effects. *See* ALUMINUM ALLOYS; CRYSTAL DEFECTS; DIFFUSION; HIGH-TEMPERATURE MATERIALS; METAL, MECHANICAL PROPERTIES OF; PLASTICITY; RHEOLOGY.

William D. Nix; Jeffery C. Gibeling; Kevin J. Hemker

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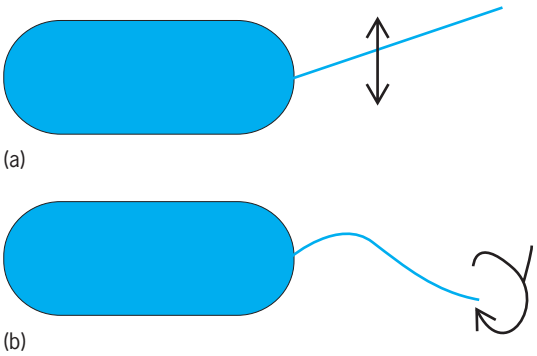
Creeping flow

Flow in which the Reynolds number is very small. Fluid dynamics encompasses an intriguing array of phenomenologically rich processes, from weather prediction in the atmosphere to the swimming of the tiniest microorganism in its natural environment. A qualitative measure of the expected behavior in a given fluid dynamics system is found from the Reynolds number, Re = density · length · velocity/viscosity. In the case where Re is very small (much less than 1, usually), the inertia effects (momentum, acceleration, and body forces) become negligible in comparison to the viscous resistance. This situation is described as creeping flow. *See* REYNOLDS NUMBER; VISCOSITY.

Creeping flows exist typically when the fluids involved are very viscous (oil, syrup) or the length scales involved are very small (swimming microorganisms, microelectronic machines, flows in cracks or pores, the transport of sediment or pollutants). In many cases of creeping flow, it is possible to significantly simplify the governing Navier-Stokes equations, which are highly nonlinear, by setting the Reynolds number identically to zero. Under this simplification, a set of equations is obtained called the Stokes equations, which have the decided advantage of being linear. As a result, it is possible to draw upon the principle of linear superposition as well as complex variable theory in search of mathematical solutions. *See* COMPLEX NUMBERS AND COMPLEX VARIABLES; LINEARITY; NAVIER-STOKES EQUATIONS; SUPERPOSITION PRINCIPLE.

An example of one creeping flow solution is the flow of a uniform velocity field $\mathbf{u} = (U, 0, 0)$ past a stationary sphere of radius a. The Stokes stream function ψ is given by Eq. (1), where (r, θ) are the

$$\psi(r, \theta) = \frac{1}{2} U \left(\frac{a^3}{2r} - \frac{3ar}{2} + r^2 \right) \sin^2 \theta \qquad (1)$$



Aquatic propulsion by (*a*) a mechanical fish and (*b*) a swimming microorganism.

polar coordinates about the sphere and lines of constant values of ψ are stream lines for the flow. From this solution, the drag D acting on this sphere can be shown to be given by Eq. (2), where μ is the fluid

$$D = 6\pi\mu Ua \tag{2}$$

viscosity. For small values of Re, inertial effects can be incorporated through a small correction to this drag formula, known as the Oseen correction. *See* STREAM FUNCTION.

Another consequence of linearity is reversibility: streamlines remain unchanged when the direction of motion is reversed. This has profound implications for aquatic animal propulsion in creeping flow. For example, a mechanical "fish" that has a tail which flaps back and forth (**illus.** *a*) may swim quite readily in the bath tub, but will make no progress in a very viscous fluid such as corn syrup. Since most swimming microorganisms exist in a low-Reynolds-number environment, they must adopt a non-time-reversible motion, such as having a corkscrew-type tail (illus. *b*). *See* FLUID FLOW; FLUID-FLOW PRINCIPLES; FLUID MECHANICS. Mary-Catherine Kropinski

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Creodonta

An extinct order of mammals containing the dominant carnivores in North America, Europe, Asia, and Africa during much of the Cenozoic. Its members roamed the Earth for more than 50 million years, then became extinct less than 9 million years ago. Creodonts were not part of the order Carnivora, but they independently evolved carnivorous specializations in their teeth and limbs. They ranged from tiny (*Isohyaenodon* from Africa was the size of a small weasel) to gigantic (*Hyainailouros* from Europe and Asia was one of the largest land-dwelling mammalian carnivores ever). The wolflike genus *Hyaenodon* is probably the most familiar creodont among the more than 180 known species. Most lived in North America and Europe, but new species from Africa and Asia are being described at an incredible rate. *See* DENTITION.

There are two major families within Creodonta: Hyaenodontidae and Oxyaenidae. Hyaenodontids are the more diverse, with more than 140 known species. They are first recognized from the mid-Paleocene of eastern Asia. They had emigrated to North America and Europe by the earliest Eocene. Their greatest diversity was during the latest Eocene and Oligocene, when more than 30 species of the genus *Hyaenodon* (see **illus.**) were doglike predators, feeding on the variety of hoofed, plains-dwelling mammals in North America, Asia, and Europe. The dentition of *Hyaenodon* was catlike, and it has been estimated that the diet of these animals was 70–100% meat. Their bodies, however, were wolflike, and they may have hunted in a manner more common to canids than felids. Early hyaenodontids were more diverse: *Prolimnocyon* was an arboreal, genet-like animal, *Prototomus* was a generalized, raccoon-like creature, and *Pterodon* was large and bearlike. Some hyaenodontids were carnivores (*Prototomus*), others were probably omnivores (*Masrasector*), some were insectivorous (*Thereutherium*), and some may have been mollusk eaters (*Teratodon* and *Apterodon*). The range of locomotory and dietary adaptations in Hyaenodontidae was probably almost as great as in Carnivora, although there were no fully aquatic creodonts (*Apterodon* from Africa may have been partially aquatic). The last hyaenodontid, *Dissopsalis*, lived in southern Asia and eastern Africa and became extinct only about 8 million years ago. Oxyaenids were less diverse (only about 40 species are known) and reached their peak in the latest Paleocene and earliest Eocene. By the late Eocene, the family had become completely extinct. While oxyaenids were specialized, the range of niches they occupied was not as great as hyaenodontids. Many oxyaenids, such as *Oxyaena* and *Dipsalidictis*, were like large wolverines, although the largest, *Sarkastodon* from Mongolia, was almost bearlike. There were a few saber-toothed oxyaenids, including *Apataelurus* from the mid-Eocene of North America; unfortunately, little is known about these.

The formal classification of Creodonta has changed considerably, largely because the relationships among the various creodont groups and other placental mammals are still poorly understood. In the 1950s, the order included many mammalian groups that are now considered part of Carnivora (true carnivores), Lipotyphla (insectivores), or Ungulata (hoofed mammals). Creodonts later formed the core of the order Deltatheridia; however, that grouping is no longer used, largely because the name-bearing genus *Deltatheridium* is now known to be related to marsupials. Improved knowledge about the relationships of placental mammals has resulted in many former creodont families being moved to other orders

so that only oxyaenids and hyaenodontids remain. It is unclear how Hyaenodontidae and Oxyaenidae are related to one another, and their grouping as Creodonta may turn out to be unjustified.

It was long thought that the extinction of the creodonts was caused by competition from better-adapted members of the order Carnivora. However, the timing of creodont evolution does not support this theory. It appears that carnivores evolved before creodonts, the former being first known from the earliest Paleocene, while the latter first evolved in the mid to late Paleocene. Creodonts and carnivores diversified side by side through the Eocene, apparently playing different ecological roles. The persistence of creodonts until the end of the Miocene (more than 50 million years of coexistence with Carnivora) also belies the competitive scenario. Furthermore, studies of creodont adaptations do not support the long-touted idea that they were inferior to contemporary true carnivores. It is more likely that chance and changing conditions, including the evolution and extinction of creodont prey species, explain their slow demise. *See* CARNIVORA; MAMMALIA. P. David Polly

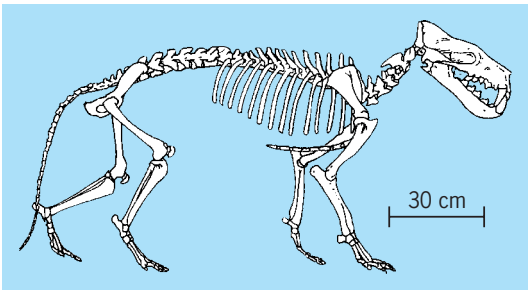
Bibliography. R. L. Carroll, *Vertebrate Paleontology and Evolution*, 1988; J. Mellett, The Paleobiology of North American Hyaenodon, *Contrib. Vert. Evol.*, 1:1–134, 1977; R. J. G. Savage and M. R. Long, *Mammal Evolution: An Illustrated Guide*, 1986.

Cress

A prostrate hardy perennial crucifer of European origin belonging to the plant order Capparales. Watercress (*Nasturtium officinale*) is generally grown in flooded soil beds and used for salads and garnishing. Propagation is by seed or stem cuttings. High soil moisture is necessary. Leafy stems are cut usually 180 days after planting. Virginia is an important producing state.

Garden cress (*Lepidium sativum*) is a cool-season annual crucifer of western Asian origin grown for its flavorful leaves. Propagation is by seed, and leaves are harvested approximately 2 months after planting.

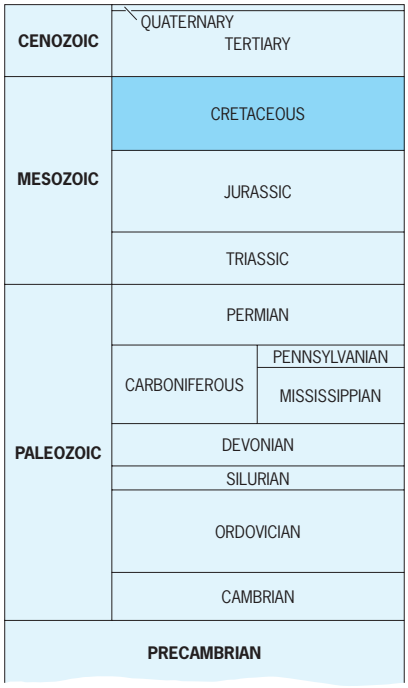
Upland or spring cress (*Barbarea verna*) is a biennial crucifer of European origin, grown and harvested similarly to garden cress but of lesser commercial importance. *See* CAPPARALES. H. John Carew



The creodont *Hyaenodon*. (*After A. S. Romer, Vertebrate Paleontology, 3d ed., University of Chicago Press, 1966*)

Cretaceous

In geological time, the last period of the Mesozoic Era, preceded by the Jurassic Period and followed by the Tertiary Period. The rocks formed during Cretaceous time constitute the Cretaceous System. Omalius d'Halloys first recognized the widespread chalks of Europe as a stratigraphic unit. W. O. Conybeare and W. Phillips (1822) formally established the period, noting that whereas chalks were remarkably widespread deposits at this time, the Cretaceous System includes rocks of all sorts and its ultimate basis for recognition must lie in its fossil remains. *See* CHALK; FOSSIL; JURASSIC; ROCK AGE DETERMINATION; STRATIGRAPHY; TERTIARY.



The parts of the Earth's crust that date from Cretaceous time include three components: a large part of the ocean floor, formed by lateral accretion; sediments and extrusive volcanic rocks that accumulated in vertical succession on the ocean floor and on the continents; and intrusive igneous rocks such as the granitic batholiths that invaded the crust of the continents from below or melted it in situ. The sedimentary accumulations contain, in fossils, the record of Cretaceous life. The plutonic and volcanic rocks are the chief source of radiometric data from which actual ages can be estimated, and suggest that the Cretaceous Period extended from 144 million years to 65 ± 0.5 million years before present (**Fig. 1**).

Subdivisions and time markers. Changes in life, by the development of new species and by the loss of old, are not uniform but somewhat steplike. Larger steps led to the recognition of periods such as the Cretaceous, while lesser steps within the Cretaceous delineate 12 globally recognized subdivisions, or stages (Fig. 1). In marine

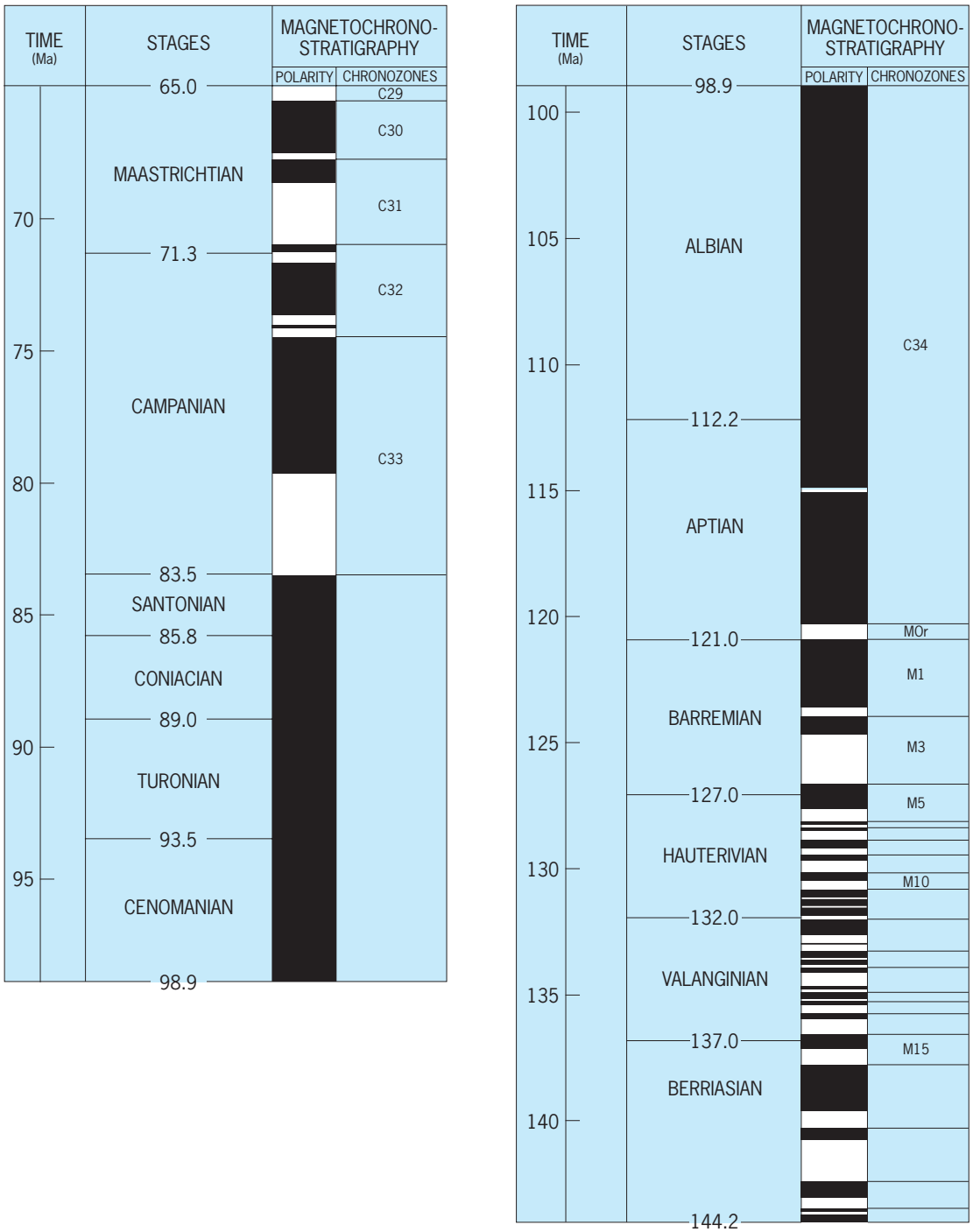


Fig. 1. Stages of the Cretaceous Period, their estimated ages in years before present, and polarity chrons representing the alternation between episodes of normal (black) and reversed (white) orientations of the Earth's magnetic field. (*After F. M. Gradstein et al., in W. A. Berggren et al., eds., Geochronology, Time Scales and Global Stratigraphic Correlations, SEPM Spec. Publ., no. 54, 1995*)

sediments the appearance and disappearance of individual, widely distributed species allows further time resolution by so-called zones. Initially these zones were largely based on ammonites, a species-rich and rapidly evolving, now extinct group of cephalopods, closest to the squids and octopuses but resembling the pearly nautilus in possession of a coiled calcareous hydrostatic shell. Over 100 successive ammonite zones have been recognized, but due to the provinciality of some species and the rarity of many, the dating of Cretaceous marine sediments now mainly devolves on microscopic fossils of the calcareous plankton. The most important of these are protozoans—the infusorial (tintinnid) calpionellids and the sarcodine planktonic foraminiferans. Their shells, in the range of 0.1–1 mm, occurring by hundreds if not thousands in a handful of chalk, have furnished about 38 pantropical zones. Next in importance are the even tinier (0.01-mm) armor plates of "nanoplanktonic" coccolithophores—

prymnesiophyte algae, best studied by electron microscopy. Thousands may be present in a pinch of chalk, and 24 zones have been recognized. *See* CEPHALOPODA; COCCOLITHOPHORIDA; FORAMINIFERIDA; MARINE SEDIMENTS; MICROPALEONTOLOGY; PALEONTOLOGY; PHYTOPLANKTON; PROTOZOA; ZOOPLANKTON.

The Earth's magnetic field reversed about 60 times during Cretaceous time, and the resulting polarity chrons have been recorded in the remanent magnetism of many rock types (Fig. 1). The actual process of reversal occurs in a few thousand years and affects the entire Earth simultaneously, providing geologically instantaneous time signals by which the continental and volcanic records can be linked to marine sequences and their fossil zonation. Noteworthy here is the occurrence of a very long (32-million-year) interval during which the field remained in normal polarity. *See* PALEOMAGNETISM.

Geography. During Cretaceous time the breakup of Gondwana, the great late Paleozoic-Triassic supercontinent, became complete (**Fig. 2**). For pre-Cretaceous time the positions of continents can be roughly derived from the remanent magnetic vectors in their rocks, which record the directions toward and the distance to the poles. But for Cretaceous and later times the positions may be more precisely tracked by the "growth lines" of crustal plates, visible in magnetic maps of the oceanic areas. As the Earth's magnetic field reversed, the continuous growth of ocean floors along the mid-oceanic ridge system occurred alternately in normal and in reversed polarity. Imprinted in the rocks, the remanent magnetism reinforces the present magnetic field above the crust grown in normal polarity and diminishes it over that grown in reversed fields.

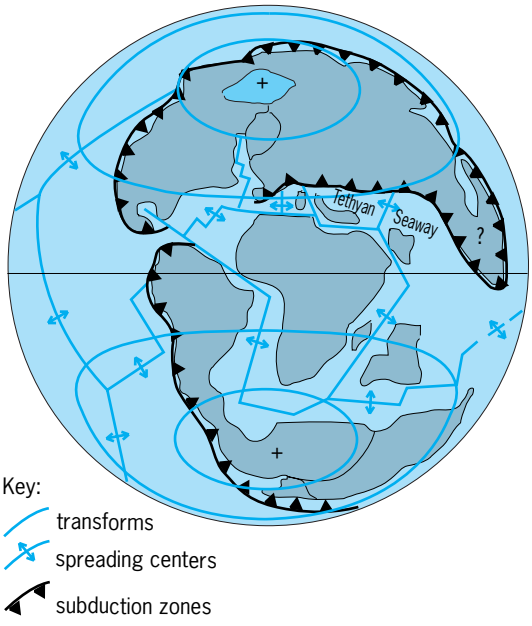


Fig. 2. Late Cretaceous global tectonics, showing inferred lithosphere plate configurations with spreading, subduction, and transform boundaries; generalized continental outlines are shown for reference. (*After R. H. Dott, Jr., and R. L. Batten, Evolution of the Earth, 3d ed., McGraw-Hill, 1988*)

Laurasia had already separated from Africa by the development of Tethys and became split into North America and Eurasia by the opening of the North Atlantic (though tenuous land connections continued to exist in the far north). These new, deep oceanic areas continued to grow in Cretaceous time. India broke away from Australia and Australia from Antarctica. South America tore away from Africa by the development of the South Atlantic Ocean, while India brushed past Madagascar on its way north to collide with southeast Asia. As these new oceanic areas grew, comparable areas of old ocean floor plunged into the mantle in subduction zones such as those that still ring the Pacific Ocean, marked by deep oceanic trenches and by the development of mountain belts and volcanism on adjacent continental margins (**Fig. 3**). *See* CONTINENTS, EVOLUTION OF; PLATE TECTONICS; SUBDUCTION ZONES.

The face of the globe was also affected by changes in sea level. Sea level at times in the early Cretaceous stood at levels comparable to the present, but subsequently the continents were flooded with relatively shallow seas to an extent probably not attained since Ordovician-Silurian times. Maximal flooding, in the Turonian Stage, inundated at least 40% of present land area. Cretaceous seas covered most of western Europe, though old mountain belts such as the Caledonides of Scandinavia and Scotland remained dry and archipelagos began to emerge in the Alpine belt. In America (Fig. 3), seas flooded the southeastern flank of the Appalachian Mountains, extended deep into what is now the Mississippi Valley, and advanced along the foredeep east of the rising Western Cordillera to link at times the Gulf of Mexico with the Arctic Ocean. In the far west, accretionary prisms of subducted Cretaceous deep-water fans and oceanic sediments, partly mixed with ophiolites in the Franciscan melange, became juxtaposed with forearc sediments.

Large seas extended over parts of Asia, Africa, South America, and Australia. The wide spread of the chalk facies is essentially due to this deep inundation of continents, combined with the trapping of detrital sediments near their mountain-belt sources, in deltas or in turbidite-fed deep-water fans such as the classical Alpine Flysch. At the same time, carbonate platforms were still widespread, and the paratropical dry belts were commonly associated with evaporite deposits. *See* PALEOGEOGRAPHY; SALINE EVAPORITES.

Tectonic and igneous activity. The North American plate, in the process of separating from Europe, continued to have a western convergent margin of growing mountains, and a southeastern passive margin. Convergent or active margins such as those surrounding the Pacific Ocean (then as now a "ring of fire") became intruded by great batholiths of granitic composition, such as those of the British Columbia Coast Range, the Sierra Nevada, the Peninsular Ranges of southern and Baja California, and the great Andean batholiths. Above these there accumulated superstructures of andesitic volcanics. Exotic

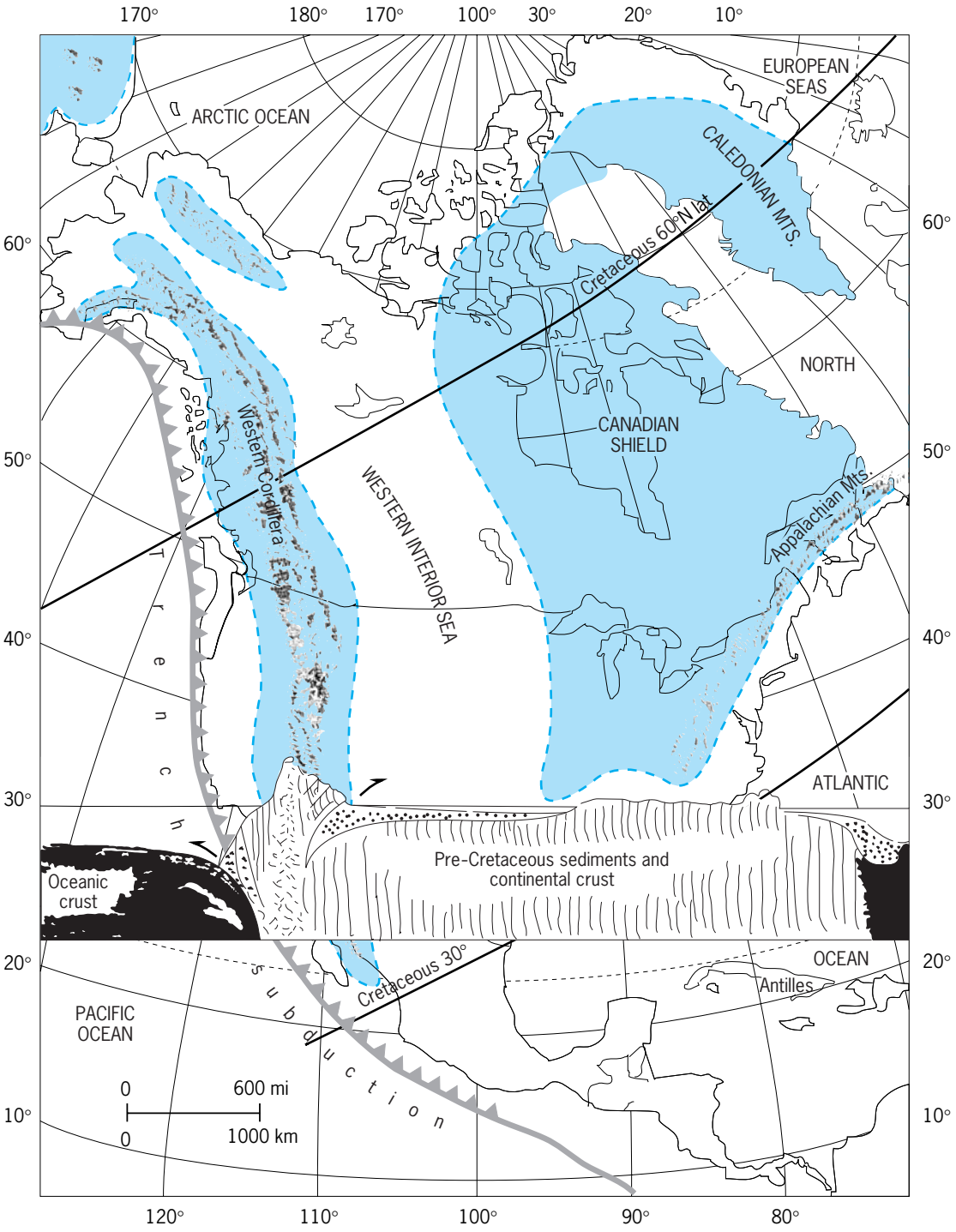


Fig. 3. Schematic view of North America in Cretaceous time. Cretaceous seas are shaded, and Cretaceous sediments in cross section (in the black-and-white strip) are dotted. Note the counterclockwise rotation of the continent shown by the lines of Cretaceous latitude.

island arc terranes riding on the Pacific plate but resisting subduction came to be welded to this margin of North America. The eastern margin of this Western Cordillera overthrust and incorporated the margin of the subsiding Western Interior Foredeep, presaging development of the Rocky Mountains (Fig. 3). Passive continental margins, such as those bordering the Atlantic, came to be sites of massive sediment accumulations, the classical geosynclines of older literature, from which mountain systems have developed locally as in the Caribbean or have yet to arise. *See* GEOSYNCLINE.

Flood basalts of Cretaceous time include the early Aptian outpourings of basalts on the mid-Pacific floor, perhaps the largest on record. Early Cretaceous emplacement of the Rajmahal basalts of India and Bangladesh was followed by the eruption of the Deccan basalts of India, emplaced during the

short magnetic interval (chron 29R) that includes the Cretaceous/Tertiary (K-T) boundary. The beginnings of the "Brito-Arctic" basalts also go back to Maastrichtian time.

Climate and oceans. In parts of early Cretaceous time, ice extended to sea level in the polar regions, as shown by glendonites, concretionary calcite pseudomorphs of the cold-water mineral ikaite; and in marine mudstones containing exotic pebbles, dropstones melted out of icebergs. But during most of Cretaceous time, climates were in the hothouse or greenhouse mode, showing lower latitudinal temperature gradients. Tropical climates may have been much like present ones, and paratropical deserts existed as they do now, but terrestrial floras and faunas suggest that nearly frost-free climates extended to the polar circles as did abundant rainfall, and no ice sheets appear to have reached sea level.

The cause for these climatic changes may be sought in variations in the heat-retaining character of the atmosphere (abundance of greenhouse gases such as carbon dioxide or methane) or in oscillations of solar luminosity. The hydrologic cycle appears to have been intensified, with more heating and therefore more water evaporation in the tropics, leading to massive transport of water vapor to the polar regions warmed by its condensation.

With a mean temperature of about 38°F (3°C), the present ocean lies 27°F (12°C) below the mean surface temperature of the Earth. This refrigerator is maintained by the sinking of cold waters in the circumpolar regions. Temperatures calculated from the oxygen isotope ratios in Cretaceous deep-water foraminiferans yield values as high as 62°F (16°C), and suggest an ocean much closer to mean surface temperature. In this ocean circulation, visualized long ago by T. C. Chamberlin, warm saline waters of the dry paratropics were the main source of dense waters and sank to the bottom when only moderately cooled. The implication is not only altered current directions and deep-water temperature regimes but also a diminished oxygen supply to the depths. This diminished oxygen is documented by the widespread development of black shales, deposited on bottoms from which most or all scavengers were excluded by lack of oxygen. The resulting excess burial of organic matter yielded an abundance of petroleum source beds, reflected in the large petroleum reserves in Jurassic and Cretaceous rocks. Occasional episodes of black shale deposition, such as that of the Bonarelli event near the end of the Cenomanian Stage, became circumglobal and buried enough organic matter to alter the carbon isotope ratios in the entire ocean-atmosphere system, but the circumstances of their origin remain unknown. *See* OCEANOGRAPHY; OIL SHALE; PALEOCLIMATOLOGY; PETROLEUM RESERVES.

Life. A walk along a Cretaceous beach or a visit to a surf-beaten cliff would have yielded many snail and bivalve shells and sea urchins differing from living ones only at the level of species. Yet there are some notable differences. Two families of bivalves rose to particular prominence in Cretaceous time. The inoceramids were a widespread and diversified group that developed species 5 ft (1.5 m) wide, the largest of all known clams, while the rudistids came to resemble corals in form and largely replaced corals as reef builders in Cretaceous time. Among cephalopods, ammonites were prominent swimmers in Cretaceous seas, as were belemnites—squids with a heavy calcareous skeleton. The calcareous microplankton had only in latest Jurassic time reached an abundance sufficient to produce widespread chalks, and in the Cretaceous two major elements of this, the prymnesiophyte coccolithophorids and the planktonic foraminiferans, reach maximal species diversity. Bony fishes evolved extensively in Cretaceous time, but prominent among marine predators were reptiles, some of them gigantic. Marine turtles have persisted, the dolphinlike ichthyosaurs died out in Turonian-Coniacian time, while the massive, sea-lion-like plesiosaurs and the lizard-derived mosasaurs lasted to the end of the period. Flightless birds populated some of the late Cretaceous seas. *See* BIVALVIA; GASTROPODA; REPTILIA.

On land, flowering plants (angiosperms) first appeared in early Cretaceous time, as opportunistic plants in marginal settings, and then spread to the understory of woodlands, replacing cycads and ferns. In late Cretaceous time, evergreen angiosperms, including palms, thus came to dominate the tropical rainforests. Evergreen conifers maintained dominance in the drier midlatitude settings, while in the moist higher latitudes forests of broad-leaved deciduous trees dominated. Grasslands, however, were not developed until Tertiary time. Insects became highly diverse, and many modern families have their roots in the Cretaceous. Amphibians and small reptiles were present. Larger land animals included crocodiles and crocodilelike reptiles, turtles, and dinosaurs. Dinosaurs, derived from reptilian stock in Triassic time, rose to become the largest land animals of all time. The very size of these animals implies complex circulatory, digestive, and respiratory systems unlike those of reptiles. Evidence for temperature control is emerging, and it is speculated that some dinosaurs were feathered. Thus, there is growing support for classification of dinosaurs not as a branch of the class Reptilia but as a separate class. The mammals, another reptilian offshoot in the Triassic, remained comparatively minor elements in the Cretaceous faunas. In early Cretaceous time, egg-laying and marsupial mammals were joined by placentals, but Cretaceous mammals were in general small, and lack of color vision in most modern mammalians suggests a nocturnal ancestry and a furtive existence in a dinosaurian world. Birds had arisen, from dinosaurs in Jurassic time, but their fossil record from the Cretaceous is poor and largely one of water birds. More common are the remains of flying reptiles, the pterosaurs, which in Cretaceous time reached a wingspan of 35 ft (11 m), the largest known flying animals by far. *See* DINOSAURIA; MAMMALIA; MARSUPIALIA; PTEROSAURIA.

K/T boundary crisis. Most species of the inoceramid and rudistid bivalves died out in a crisis within the Maastrichtian Stage. Some millions of years later,

during the reversed magnetic interval known as chron 29R, a collision with an asteroid or comet showered the entire Earth with impact debris, preserved in many places as a thin "boundary clay" enriched in the trace element iridium. This event coincided with the great wave of extinctions—the K/T crisis—which serve to bound the Cretaceous (Kreide) Period against the Tertiary.

The Chicxulub crater is located at the northwestern edge of what was then the Yucatán Carbonate Platform and now the Yucatán Peninsula. Buried under some thousands of feet of Tertiary limestone, it forms a geophysical anomaly explored for oil by a number of boreholes. With a diameter of 110 mi (180 km) excluding possible peripheral rings, this is the largest crater known from Phanerozoic time. Like comparable craters on the Moon and Mars, it has a central peak and peripheral ejecta blanket. It is filled with partly melted impact debris (suevite) derived from sediments and the underlying igneous-metamorphic complex. Quartz grains with shock lamellae confirm the impact origin, and abundant glass yields the K/T boundary age of 65 $\pm$ 0.5 million years. The ejecta blanket, where exposed at the surface in the Mexico-Belize border country 300 mi (400 km) distant, consists mainly of middle to late Cretaceous carbonates in blocks up to 10 ft (3 m), mixed with carbonate lapilli and altered glass lumps. Rare cobbles of limestone and chert were melted and quenched. In surrounding regions, deeper-water marls of latest Cretaceous and earliest Tertiary age are separated by a few feet of unusual sediment: a thin bed contains chips of limestone, blebs of glass generally altered to clay, and shocked quartz. In the Gulf of Mexico region this is followed by cross-beds interpreted as deposits of a "tidal wave" (tsunami) that swept repeatedly across the Gulf. A centimeters-scale layer of clay, extending around the world, contains microscopic silicates, while electron microscopy reveals crystals of spinel with excessive degrees of oxidation. Shocked quartz, abundant and relatively coarse in America, becomes scarce and fine with distance.

The iridium content associated with the fallout implies an impactor with a composition similar to that of planetary interiors. The size, estimated at 6 mi (10 km), implies an asteroid or a correspondingly larger comet. Traveling at 16 mi/s (25 km/s), the body would have struck with the explosive force of 10^{14} tons of TNT. Quite aside from local and regional devastation, global effects must have included earthquake shock many orders of magnitude greater than any found in human history; associated land slips and tidal waves; a dust blackout of sunlight that must have taken many months to clear; a sharp drop in temperatures that would have brought frost to the tropics; changes in atmospheric and water chemistry; and disturbance of existing patterns of atmospheric and oceanic circulation.

It is possible that earthquakes influenced volcanic eruptions. The first flows of the Deccan basalts predate polarity chron 29R and the impact, but the great body of basalt poured out within that brief chron, possibly emitting enough sulfur to add to the crisis.

Different biotic communities were affected to different degrees. The pelagic community, sensitive to photosynthetic productivity, was severely struck, with coccolithophores and planktonic foraminiferans reduced to a few species, while ammonites, belemnites, plesiosaurs, and mosasaurs were eliminated. Yet dinoflagellates, endowed with the capacity to encyst under stress, suffered no great loss. Benthic life was only moderately damaged beyond the loss of species, excepting destruction of the reef community. While North American trees underwent far more extinction at the specific level than formerly believed, land floras escaped with little damage, presumably because they were generally equipped to handle stress by dormancy and seed survival. The plant-fodder-dependent dinosaurs perished, as did their predators and scavengers. The fresh-water community, buffered by ground water against temperature change and food-dependent mainly on terrestrial detritus, was little affected. While a great many individual organisms must have been killed by the immediate effects of the impact, the loss of species and higher taxa must have occurred on land and in shallowest waters mainly in the aftermath of darkness, chill, and starvation, and in the deeper waters in response to changed regimes in currents, temperatures, and nutrition.

The Cretaceous crash led above all to an evolutionary outburst of the mammals, which in the succeeding tens of millions of years not only filled and multiplied the niches left by dinosaurs but also invaded the seas, to become successors to plesiosaurs and mosasaurs. Humans, among others, owe their existence to that devastating event of 65 million years ago.

Alfred G. Fischer

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Criminalistics

Criminalistics is the science and profession dealing with the recognition, collection, identification, individualization, and interpretation of physical evidence, and the application of the natural sciences to law-science matters. The term originated from the book *Handbuch für Untersuchungsrichter als System der Kriminalistik* (3d ed., 1898) by Hans Gross, an investigating magistrate and professor of criminology at the University of Prague. He described the need for a scientifically trained investigator who could undertake certain technical aspects of an investigation and could also serve as liaison between scientific specialists who might assist in the investigation of criminal activity. This concept was popular in Europe, where a number of forensic science institutes were developed to apply the tools and techniques of the natural sciences to the investigation

of crime and, generally, in official governmental inquiries.

In the United States and the United Kingdom, with legal systems fundamentally different from those of Europe, the criminalistics profession has developed in a different fashion. Criminalistics laboratories were established in a few police departments in the early decades of the twentieth century (in Chicago, Los Angeles, and New York) and in the Federal Bureau of Investigation (FBI). In the 1970s an infusion of money from the federal government resulted in the formation of numerous laboratories associated with state and local law enforcement agencies.

The majority of criminalists work in laboratories associated with governmental agencies charged with enforcing local, state, or federal laws or regulations. However, a number of criminalists engage in private practice or are employed by academic institutions.

Education. Criminalists require a broad education in the natural sciences, including inorganic, organic, and analytical chemistry; physics; mathematics and statistics; and biology and biochemistry. In addition, curricula in criminalistics or forensic science leading to bachelor's or master's degrees include courses dealing with specific types of commonly encountered evidence materials (such as dried biological fluid stains, impression evidence, or trace evidence), with specific analytical procedures (such as microscopy or instrumental analysis), and with applicable legal issues.

As a consequence of the small number of graduates with degrees in criminalistics or forensic science,

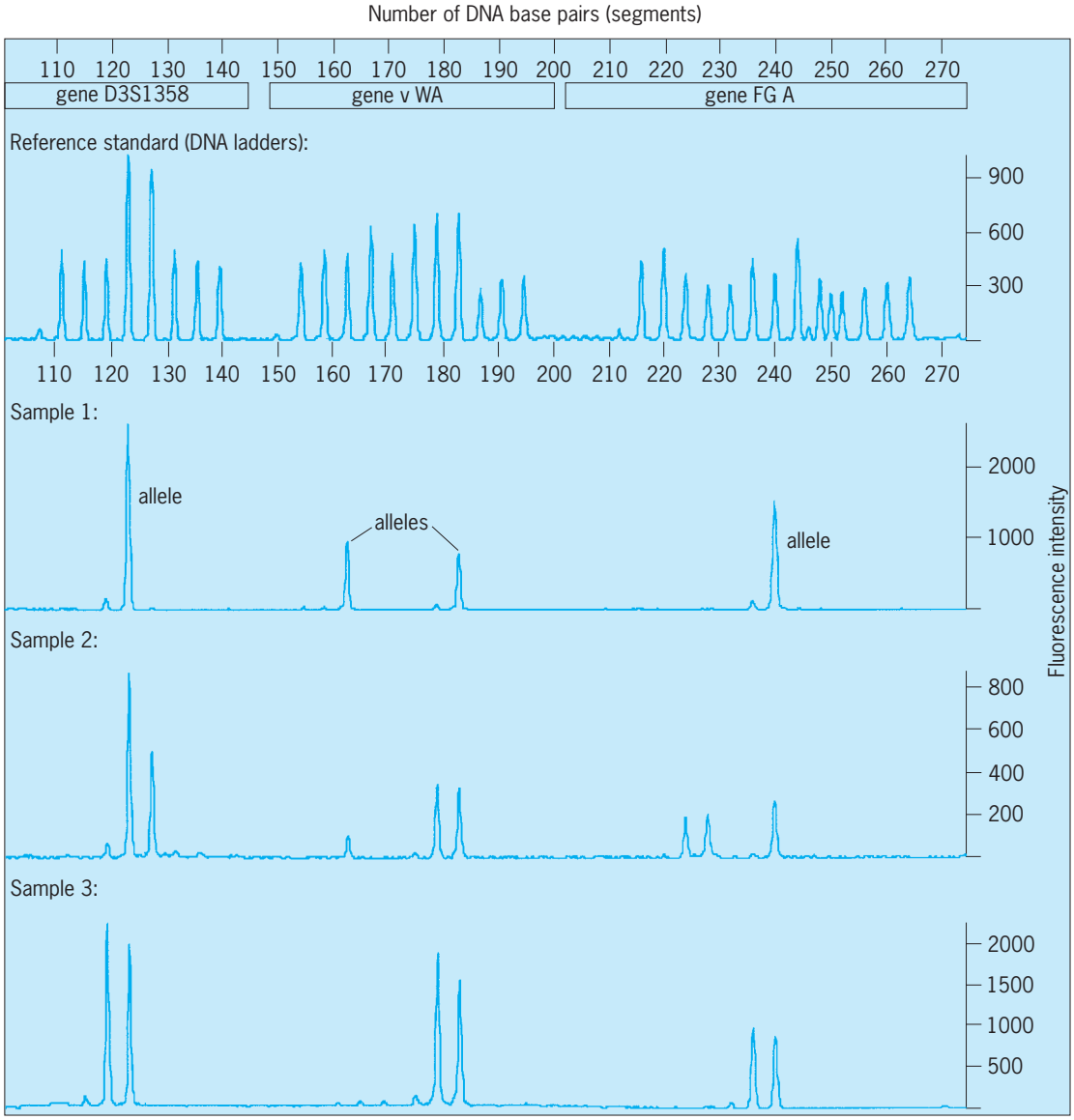

Fig. 1. Capillary electropherogram of DNA isolated from biological evidence. DNA, isolated from biological material such as blood, semen, hair, or tissue, can be compared with DNA obtained from suspects or victims of crimes. The polymerase chain reaction (PCR) is used to increase minute amounts of DNA isolated from evidenciary specimens. Analysis by capillary electrophoresis, or other techniques, allows the criminalist to determine, with a very high degree of certainty, that the evidenciary sample is or is not from a particular individual. (*Forensic Science Associates*)

many criminalists have undergraduate or graduate degrees in scientific disciplines such as chemistry, biochemistry, genetics, or biology, and have chosen a career in criminalistics after completing their education. A wide variety of in-service training is available, including training through the FBI, the California Department of Justice (through the California Criminalistics Institute), and various professional associations.

Specialists. It is not possible for a single person to become proficient in the examination and analysis of all types of physical evidence. Increasingly, criminalists and other workers in forensic science laboratories are specializing in the examination of one or a few types of physical evidence.

Forensic biologists. Forensic biology is the analysis of the biological or genetic properties of evidence. Criminalists specializing in forensic biology are involved in the identification of biological evidence and attempts to determine its source. Traditional methods for the analysis of blood group antigens and genetically variant proteins present in blood, semen, and other biological materials have given way to the analysis of DNA (deoxyribonucleic acid) which can be recovered from such materials. (**Fig. 1**). The ability to identify the individual from whom a blood or semen sample has come has revolutionized the science of criminalistics. *See* DEOXYRIBONUCLEIC ACID (DNA); FORENSIC MEDICINE.

Trace evidence analysts. These specialists analyze material that is transferred between two objects which come into contact (**Fig. 2**). The Locard exchange principle, credited to the French criminalist Edmond Locard, states that whenever two objects come into contact, portions (or traces) of one object will be transferred to the other. Finding these traces, identifying what they are, and determining their origin through the process of individualization (determination of the parameters of a sample which will allow it to be distinguished from other or all similar items) is the job of the trace evidence analyst. This analysis is based on the chemical or physical properties of the material. Use of the optical microscope and the electron microscope and highly sensitive methods of chemical analysis (such as Fourier transform infrared spectroscopy, energy dispersive x-ray analysis, mass spectroscopy, and neutron activation analysis) are typically used by trace evidence specialists. *See* MASS SPECTROSCOPE; SCANNING ELECTRON MICROSCOPE; SPECTROSCOPY; X-RAY FLUORESCENCE ANALYSIS.

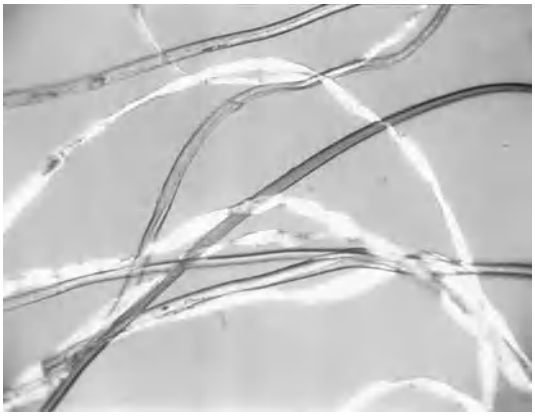

Fig. 2. Fiber photomicrographs. Trace evidence such as fibers, hairs, soil, glass, paint, or botanical material may be transferred between two objects which come into contact. A variety of analytical techniques, including polarized light microscopy, fluorescence microscopy, infrared spectroscopy, scanning electron microscopy, energy-dispersive x-ray spectroscopy, and microspectrophotometry are used for the analysis of such materials. The finding of material from one object on another object provides evidence that the two objects were in contact. (*Forensic Science Associates*)

Firearms and toolmark examiners. These specialists examine firearms, ammunition components, and tools and marks left by them. The underlying principle is that when a tool acts on some object it will leave a mark which is unique due to the configuration of the cutting edge. The uniqueness of each tool is a result of manufacturing processes and postmanufacturing wear and damage. The tool may be a screwdriver which was used to pry open a door or window, or a gun barrel which produced distinctive markings on a fired bullet. The firearms and toolmark specialist will compare the marks on a recovered bullet or a toolmark on a window with guns or tools recovered during the investigation. It is often possible to conclude that a recovered bullet was fired from a particular firearm or that a mark on a window was made by a particular screwdriver.

Other evidence specialists. At a crime scene, physical evidence is routinely encountered (for example, shoe or tire impressions, gunpowder residues on the body or clothing of victims or perpetrators in shootings, fragments of bombs or destructive devices recovered from scenes of sabotage, or pieces of botanical material). There are people who specialize in some of these areas, but often a criminalist is called upon to develop a method for examination of unique evidence, or to consult with a scientist from an industrial or academic laboratory with expertise in an unusual field (such as forensic anthropology).

Criminalists are often involved in the analysis of suspected illegal drugs and narcotics, the examination of questioned documents, or fingerprint identification. Although they are familiar with the techniques used for the examination of these types of evidence, the examinations are usually performed by specialists. *See* FINGERPRINT; FORENSIC CHEMISTRY; FORENSIC TOXICOLOGY.

Analytical techniques. A wide variety of techniques are used by criminalists for the location and collection of evidence at crime scenes as well as for the examination and analysis of that evidence in the laboratory. Crime scene techniques may involve the use of lasers or other light sources to locate biological stains or minute fibers or paint particles, chemical tests for lead around suspected bullet holes, electrostatic devices to recover a dusty shoe sole impression from a floor, or special reagents for the development of latent fingerprints.

Many techniques used in the forensic laboratory are the same ones that are used by analytical

chemists, molecular biologists, materials scientists, and so on. Often these techniques are adapted to the special requirements of the forensic science laboratory. Infrared spectroscopy, mass spectrometry, gas chromatography, optical and electron microscopy, and a host of other standard analytical chemistry techniques find common use by criminalists. *See* GAS CHROMATOGRAPHY.

Routine techniques and procedures have been developed by forensic scientists which have little or no application outside the forensic laboratory. Examples are the determination of genetic markers in minute fragments of dried biological material, the determination of the refractive index of microscopic glass fragments, the microscopic comparison of individual human hairs, and the microscopic comparison of markings on the surface of bullets. Peter D. Barnett

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Crinoidea

A class of exclusively suspension-feeding echinoderms with long, slender arms arranged radially around the calyx, a rigid cuplike structure composed of calcareous plates. The radial arm arrangement gives crinoids a flowerlike appearance (**Fig. 1**). Two basic adult body types are recognized: the sea lilies, with a long, anchored stem vertically supporting the calyx and arms above the sea bottom; and the stemless featherstars, or comatulids, with a whorl of flexible appendages on the calyx. Crinoids have a worldwide distribution and can be found in all seas except the Black and Baltic. They occupy depths ranging from just below sea level to a depth of over 9000 m (29,500 ft). Sea lilies are found only at depths greater than 100 m (330 ft), whereas comatulids are most abundant and diverse in shallow, tropical coral reef environments. *See* BLASTOIDEA; PELMATOZOA.

Over 500 species of living comatulids have been described, and many of these can be extremely abundant locally, yet fewer than 100 species of living sea lilies are known. These patterns contrast starkly with the fossil record of crinoids, which is dominated by stemmed crinoids: of the more than 5000 described fossil species, more than 90% have stems.

Living crinoids have no economic importance and are not used for food by humans. However, their fossil remains are often the dominant constituent of building limestone (for example, Indiana limestone) which is highly valued.

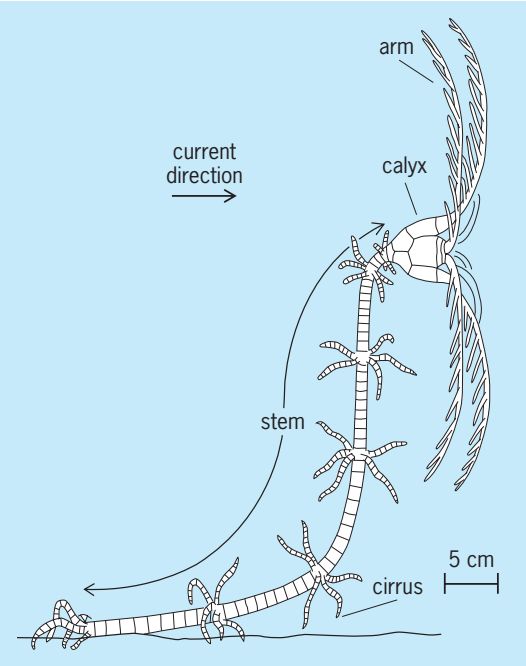

Fig. 1. Stemmed crinoid in feeding posture.

Morphology. Adult crinoids range in size from a few centimeters for some of the stemless forms to several meters from base of the stem to tip of the arms for the largest stemmed crinoids. They also vary in color from the largely bland whites and grays of the deep-water forms to the brilliant reds, yellows, and purples of the shallow-water tropical featherstars. In spite of the size range, color differences, and stemmed or stemless condition as adults, all crinoids share a suite of morphological traits that can be traced back to the Ordovician age, nearly 500 million years ago. *See* ORDOVICIAN.

As in other echinoderms, the crinoid skeleton is composed of many calcareous plates, often numbering in the thousands. Each plate is an intricate meshwork of interconnected microscopic rods and bars of calcium carbonate. The skeletal plates are joined to each other by fibers of connective tissues that penetrate and attach to the array of rods and bars. Plates can be bound to adjacent plates by ligaments, muscles, or both, and the strength and flexibility of each articulation is determined largely by the type and arrangement of the soft tissues.

The calyx consists of tightly sutured plates arranged in circlets of five with arm plates articulated to the uppermost circlet (**Fig. 2**). A flat to domelike structure, called the tegmen, forms a lid over the cuplike calyx. The tegmen may be armored with calcareous plates or, as in most living crinoids, may consist only of a leathery skin.

The arms are constructed of one or two series of plates; in arms consisting of two series of plates, adjacent plates interdigitate with each other. Each of the five arms may remain undivided, or may branch one or numerous times producing as many as 200 branches. Arm plates may bear short, lateral

branches, called pinnules, composed of serially arranged plates (**Fig. 3**).

As juveniles, all crinoids possess a stem. However, comatulids lose the stem early in ontogeny, such that in adults the lowest plate of the calyx bears only cirri, a set of long, slender articulated appendages with a clawlike terminal element used for grasping the substrate. In adult sea lilies, the stem is retained. One end of the stem articulates to the base of the calyx, and the opposite end anchors the animal to the bottom. The stem is composed of flat, disklike plates, called columnals, that are circular to pentagonal in outline, perforated at the center, and stacked plate on plate. Ligament fibers connect adjacent columnals. The anchoring end of the stem may consist of a rootlike structure that penetrates soft bottom sediments or that cements the stem to a hard substrate, or, most commonly, a set of grasping cirri. In extant sea lilies with cirri, called isocrinids, these appendages are arranged in whorls of five that are regularly spaced along the entire length of the stem. *See* ONTOGENY.

Crinoid viscera are contained in the coelom that forms the body cavity in the calyx and also extends into the arms, pinnules, stem, and cirri. The digestive system consists of a U-shaped gut with its terminal ends, the mouth and anus, penetrating the tegmen. A water vascular system, consisting of an array of fluid-filled canals in the calyx and all appendages, performs many functions associated with feeding and respiration. Its slender movable podia-like structures, called tube feet, extend along the arms and pinnules; they are sites of respiratory gas exchange and the primary food capturing organs.

Ecology and function. Crinoids are exclusively passive suspension feeders, extracting food particles from the ambient water. Food capture occurs when a microscopic particle carried by the current strikes and adheres to a mucus-coated tube foot extended from the pinnule (**Fig. 4**). Subsequently, the tube foot deposits the particle into the food groove that lines the oral side of pinnules and arms. In the food groove, other tube feet and ciliary currents transport the particle toward the mouth.

Crinoids are indiscriminate feeders, and the tube feet capture organic and inorganic particles with a median size of about 50 micrometers and rarely larger than 500 μm. The organic food component consists primarily of phytoplankton, protozoa, and crustacea.

Crinoid morphology and behavior strongly reflect their total reliance on water movement for nutrient supply. Crinoids avoid slack-water environments, living in areas dominated by currents, wave action, or multidirectional flows. Comatulids typically occupy exposed sites on reef crests, and in deep water they are often perched on corals, sponges, and sea lily stems, away from the lower velocities characterizing the boundary layer. However, some live semicryptically within the reef infrastructure. Sea lilies are found in areas dominated by currents, even in ocean depths.

Among comatulids, tube-foot morphology and

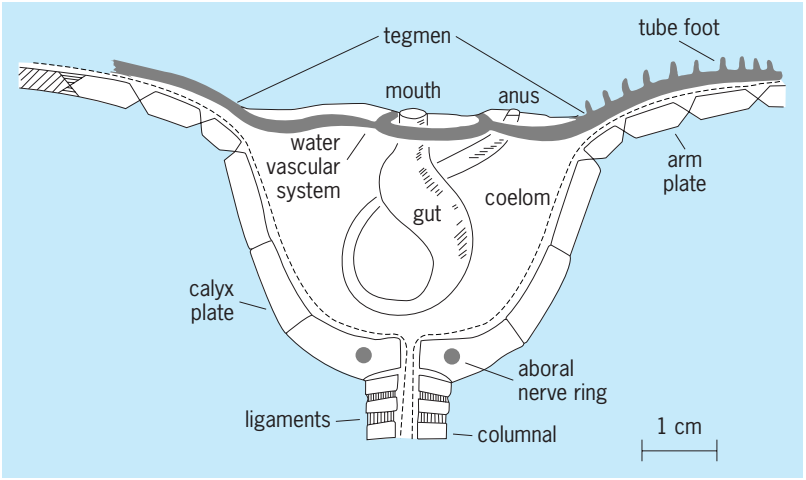

Fig. 2. Calyx and proximal stem and arm morphology, vertical section.

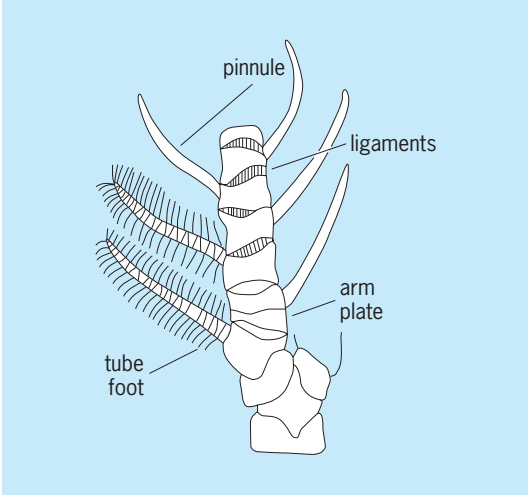

Fig. 3. Arm morphology, back (aboral) view.

spacing along pinnules varies in relation to flow velocities: species with longer, more widely spaced tube feet live in environments with lower velocities, and those with shorter, more closely spaced tube feet occupy exposed sites with higher velocities. Crinoid feeding postures vary depending on flow velocities and flow patterns. When exposed to unidirectional currents, crinoid arms, pinnules, and tube feet are aligned into a nearly planar array oriented perpendicular to flow. To avoid having the ambient current directly impinge on food grooves in which particles are transported toward the mouth, crinoids arrange all their arms with the food groove, oral side facing downstream and the opposite, aboral side facing upstream.

Under oscillating water movements produced by wave action, arms are arranged in a bidirectional fan with half the arms facing one direction and the others facing in the opposite direction. The downstream orientation of the pinnular food groove is often maintained by the weathervanelike swiveling of pinnules with their extended tube feet.

Semicryptic comatulids living within the reef infrastructure experience multidirectional flows.

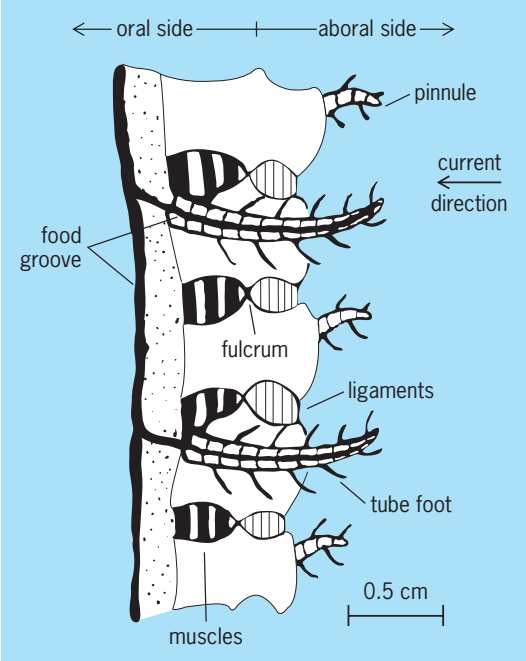

Fig. 4. Arm morphology, side view.

These crinoids do not form planar fans with their arms, but extend their arms into three-dimensional bushlike structures. Pinnules are held in two or three planes along each arm, and swiveling of pinnules occurs in response to changes in current direction.

The coordination and flexibility required for achieving the appropriate feeding posture is attained with the complex nervous system and the ligamentary and muscular arm articulations. Arm movements are generated by muscles and viscoelastic ligaments acting antagonistically across a fulcrum that lies between two adjoining arm plates. In comatulids, arm movement can be highly coordinated and rapid, allowing crawling and, in some, short bursts of swimming. Crawling also has been observed among the dominant group of living stemmed crinoids, the isocrinids, which can move slowly along the bottom by pulling with their arms and dragging their stem behind.

Some comatulids are nocturnal and use their crawling abilities to move between reef crevices and exposed sites on the reef that they occupy at night. In general, crawling allows crinoids to move to better feeding microhabitats and to avoid disturbance.

Crinoid ligaments can undergo rapid and reversible changes in stiffness, a property that is important to crinoid function and behavior. In their compliant state, ligaments allow movement of body parts with a minimum amount of muscular force. In their stiff state, ligaments can withstand high external forces, such as those due to current drag, with little deformation, allowing crinoids to maintain stable feeding postures. Through the rapid loss of stiffness and strength, ligaments can irreversibly disintegrate, allowing the animal to shed body parts. The shedding of body parts, called autotomy, can occur in response to external physicochemical stimuli, such as rapid temperature changes or predatory attacks.

Several interactions between crinoids and other animals are known to occur in the modern seas. The most common are parasitic and commensal associations between polychaete worms (Myzostomida), small snails (Eulimidae, Styliferidae), and crustaceans (Ascothoracida) that infest crinoids. These parasites can feed from the ambulacral grooves or the tegmen and can induce malformations such as cysts. Clingfishes (Gobiesocidae) are also known to feed on comatulid pinnules and on worms and copepods associated with crinoids.

Geologic history. Crinoids have a long and rich fossil record, and at times in Earth history they were numerically one of the dominant members of the benthic marine ecosystem. The first undisputed crinoids were found in rocks of the Ordovician Period, although an enigmatic fossil, *Echmatocrinus*, of Middle Cambrian age has been described as a crinoid. Crinoid diversity waxed and waned throughout the roughly 250 million years of the Paleozoic. The Mississippian Period, known as the Age of Crinoids, represents the peak in their abundance and diversity. Thick, extensive deposits consisting almost exclusively of crinoid remains, called encrinites or criquinites, are especially common in the Mississippian rock record, when crinoid diversity reached its peak of over 100 genera. Crinoids remained an important component of marine communities until the Permo-Triassic extinction event that signaled the end of the Paleozoic Era. This event led to a great reduction in crinoid diversity. A single genus, *Holocrinus*, found in Lower Triassic rocks, is the most primitive member of the Articulata, a subclass that includes all post-Paleozoic crinoids. *See* CAMBRIAN; MISSISSIPPIAN; PALEOZOIC; TRIASSIC.

Crinoids rediversified rapidly, and by the Late Triassic their range of morphologies and ecologies was as high as at any time in the Paleozoic, although the number of Mesozoic or Cenozoic genera never approached that of the Mississippian. The Mesozoic also represents an interval during which several lineages lost the stem. Some of these lineages became planktonic or pseudoplanktonic, while the comatulids, which first appeared in the latest Triassic, achieved a great level of vagility but remained in the benthos. Overall, levels of vagility not known in the Paleozoic characterize post-Paleozoic crinoids, a consequence of the highly muscularized arms and stem-shedding abilities of the latter group.

After the Mesozoic, crinoids became a much less common and less diverse part of the fossil record. Cenozoic rocks contain only bottom-dwelling crinoids, with the stemless comatulids and the cirri-bearing isocrinids most common. Whereas stemmed crinoids had been an important element of shallow-water environments throughout the Paleozoic and Mesozoic, they became restricted to bathyal and abyssal depths during the Cenozoic, which is still true today. The timing of the displacement of stemmed crinoids into deeper waters during the Cenozoic was coincident with the diversification of shell-crushing fishes. Today significant fish predation

on crinoids does not occur. The disappearance of stemmed crinoids from shallow water environments and the success there of the stemless comatulids that are highly mobile, often semicryptic, and possess certain other antipredatory features, argue for fish predation as an important ecological and evolutionary agent. *See* ARTICULATA (ECHINODERMATA); CAMERATA; CENOZOIC; CRETACEOUS; ECHINODERMATA; FLEXIBILIA; INADUNATA; JURASSIC; MESOZOIC; REGENERATION (BIOLOGY). Tomasz K. Baumiller

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Critical care medicine

The treatment of acute, life-threatening disorders, usually in intensive care units. Critical care medicine has been practiced informally for many decades in trauma centers, postanesthesia recovery rooms, coronary care units, delivery rooms, emergency rooms, and postoperative areas. The facilities and trained personnel available in the intensive care unit (ICU) permit extensive monitoring of physiological variables, organization of complex, multidisciplinary diagnostic and therapeutic plans, administration of therapy to predetermined goals, and expert nursing care.

Critical care thus runs counter to the traditional division of specialities by organ or organ system. Specialists in critical care undergo training beyond a primary qualification (internal medicine, surgery, anesthesia, or pediatrics), and must be able to manage acute respiratory, cardiovascular, metabolic, cerebral, and renal problems, as well as infections. The patients may be newborns, children, or adults suffering from trauma or acute life-threatening disease. Patients having failure of multiple organs, complicated medical problems, disorders falling into several medical specialities, or a need for 24-h care often become the responsibility of the critical care specialist.

The intensive care unit is the most labor-intensive, technically complex, and expensive part of hospital care. Of the average daily cost of a bed, about 70% is for personnel. The intensive care unit, however, may be crucial to the patient's survival.

Historical background. Although there were probably precedents, it was during the Crimean War in 1856 that Florence Nightingale placed the most serious casualties nearest the nursing station for the sake of better and more efficient care. Operational efficiency became important in the deployment of scarce medical resources and personnel, particularly in combat. In the civilian area, a special room was used for the postoperative care of neurosurgical patients at Johns Hopkins Hospital in 1923. In 1930 a postanesthesia recovery room was installed at Tübingen, Germany, as well as a unit that applied new knowledge and techniques to postoperative needs. Casualty facilities and trauma hospitals were developed in Europe around World War II. Shock wards were established in field hospitals to resuscitate battle casualties before and after their operations; the importance of blood transfusions, early operation, and postoperative care were recognized. Subsequently, trauma units, coronary care units, and postoperative care were recognized. Trauma units, coronary care units, and neonatal intensive care units were established in the United States to promote expediency and efficiency. In community hospitals the need for intensive care units was mainly the result of the shortage of nurses during and after World War II. It became essential to group patients according to the severity of their illness.

A historical example illustrates these developments. In the poliomyelitis epidemic of 1952 in Copenhagen, an estimated 5000 persons were afflicted; about 10% of these had respiratory paralysis from bulbar polio. Tubes placed through surgically prepared openings in the trachea and manual ventilation were instituted for patients with respiratory paralysis. To assist patients who could not breathe, medical teaching and research were suspended and the students worked day and night in shifts to ventilate patients by hand. This extraordinary effort reduced the mortality rate from 85 to 40%. This episode stimulated clinical research toward the development of positive-pressure mechanical ventilation. An early model of the Engstrom respirator proved to be reliable; through subsequent developments the iron lung, with its negative-pressure whole-body ventilation, was superseded. Mechanical ventilation was used for patients having crushed chest injuries and was employed during thoracic surgery and abdominal surgery. Respiratory and multidisciplinary intensive care units were rapidly developed for patients whose respiration was inadequate for any of various reasons.

Internists became involved in critical care when it was discovered that apparent death from heart attack could be treated by electric shock: dysrhythmias (abnormal heart rhythms) were found to be the most common cause of death from myocardial infarction in the first 12 h after the onset of symptoms. Resuscitation was found to be possible only if a well-equipped, trained team capable of inserting the tracheal tube and performing defibrillation arrived within minutes of cardiac arrest. This finding gave impetus to the development of paramedic transport of patients (prehospital emergency care) and in-hospital coronary care units.

In the late 1950s, modern emergency resuscitation developed. It was based on the observation that the expired breath of the rescuer could be used to breathe for the patient with ventilatory arrest. Mouth-to-mouth breathing and external cardiac massage led to the technique of cardiopulmonary

resuscitation (CPR) and subsequently to protocols for basic life support and advanced life support.

The introduction of cardiac catheterization, its application to the diagnosis and management of congenital and acquired heart disease, and its later application to cardiovascular monitoring have been basic to managing circulatory problems in patients in the intensive care unit. The advent of plastic catheters facilitated such procedures as intravascular blood sampling, infusion of fluids and titration of drugs, and circulatory measurements. In the early 1970s, the flow-directed, balloon-tip pulmonary artery catheter was introduced, which in combination with a thermodilution system permits estimation of cardiac output.

Although renal failure had been managed with peritoneal dialysis beginning in 1923, hemodialysis (dialysis of blood through an external membranous coil) came into wide clinical use only after World War II. Today, in addition to being used in outpatient and other inpatient settings, hemodialysis is employed in intensive care units, for example, to treat patients who have multiple organ failure.

Many advances in medicine have increased the need for critical care. Among them are heart operations, transplantation of various organs, and cancer chemotherapy.

Life-support systems. Perhaps the single most useful function of the intensive care unit is to provide life-support systems for desperately ill patients who would not survive without them. Such systems are briefly discussed below.

Mechanical ventilation. A mechanical ventilator automatically delivers oxygen-enriched air at positive pressure through an endotracheal tube. The machine may be driven by pressure (pressure-regulated) or volume (volume-regulated). Adjustments may be made to control rate of ventilation, maximum airway pressure, peak flow, ratio of time of inspiration to that of expiration, and oxygen concentration. Ventilator modes determine how much effort the patient makes and how much assistance the ventilator provides; the ventilator can entirely control ventilation. Different clinical conditions require somewhat different ventilator modes.

Cardiopulmonary resuscitation. Resuscitation for cardiopulmonary arrest must be begun at the site and time of arrest whether in the intensive care unit, elsewhere in the hospital, or in the community. Arrest may occur more frequently in the intensive care unit than elsewhere because the sickest patients are usually there. Out in the field, without equipment, cardiopulmonary resuscitation consists of (a) airway control (positioning of head, inflation of lungs, manual clearing of mouth and throat), (b) breathing support [mouth-to-mouth (or nose) ventilation], and (c) circulation support (control of external hemorrhage, mechanical chest compressions if the patient is pulseless, and circulatory support with intravenous fluids if available). When equipment is available, cardiopulmonary resuscitation consists of the above plus (a) pharyngeal suctioning, tracheal intubation, or, if this is not immediately successful, creation of an opening into the trachea in the neck; (b) ventilation manually (using a bag and mask) or mechanically, with oxygen added; and (c) mechanical chest compressions, or if blood pressure and pulse do not respond immediately, open-chest massage with direct cardiac compressions.

Advanced life support consists of the use of cardiac drugs, fluids, electrocardiographic monitoring for cardiac dysrhythmias, defibrillation by electrical shock with external direct current of 200–400 joules, and management of subsequent complications and multiple organ support in an intensive care unit.

Peritoneal dialysis and hemodialysis. In the intensive care unit, peritoneal dialysis and hemodialysis are used to help treat such challenging conditions as renal failure, multiple organ failure, drug overdose, sepsis (widespread infection), posttrauma conditions, and postoperative renal failure. Blood is withdrawn through intravascular catheters and pumped through a dialysis chamber and back to the patient's vascular system. Among items removed by dialysis are the waste products, urea and creatinine; excess water, potassium, and sodium; drugs; and organic acids and other noxious metabolic end-products.

Circulatory support with intraaortic balloon pumping. The intraaortic balloon pump is the temporary device that is most widely used to assist the failing heart. It is employed when the ventricle of the heart cannot pump enough blood to meet the minimum needs of the body. The balloon, which is placed in the descending aorta, is rapidly inflated in diastole (the time of cardiac relaxation) and rapidly deflated at the onset of systole (the time of cardiac contraction). When the pump is properly synchronized with the heart, the balloon deflation produces a vacuumlike effect that helps the left ventricle to contract. The device also augments diastolic pressure, which improves blood flow and oxygen supply to the heart. Combining the mechanical assistance with the ventricle's own limited ability can permit adequate circulation.

Extracorporeal membrane oxygenation. Extracorporeal membrane oxygenation equipment is a modification of the apparatus that is routinely used to oxygenate the blood outside the body during open heart surgery. It is used mainly in near-term infants who do not respond to maximum ventilatory and medical support.

Life-sustaining therapy. Another aspect of critical care is the provision of life-sustaining therapy. Components include administration of intravenous fluids, provision of nutritional support, and control of infections.

Fluid and electrolyte administration. Fluids are given mainly to replace fluid losses and to provide circulatory support in patients with conditions such as shock or circulatory deficiencies. Intravenous and intraarterial catheters are placed to allow the monitoring of arterial and central venous pressures and the rapid administration of fluids, such as blood, blood components, and various solutions.

Drugs often are used in conjunction with fluid therapy. Such drugs include agents that stimulate cardiac contraction, agents that increase blood

pressure, and agents that reduce vascular resistance to cardiac output.

Nutritional support. In life-threatening crises, nutritional requirements must be maintained. A patient who cannot take nourishment by mouth receives it either via a tube introduced through the nose into the stomach or duodenum, or parenterally (through a vein). Glucose, fat, amino acids, and other substances are given. Total parenteral nutrition can sustain life indefinitely in patients who are unable to eat and have higher than normal nutritional requirements because of sepsis, injury, or prior malnutrition.

Control of infection. A large percentage of patients in intensive care units have infections that are part of their primary disease or were acquired in the hospital. The latter infections are particularly difficult to eradicate because they are usually resistant to most antibiotics. Infectious agents are identified by culturing bronchial secretions and the fluids from body cavities, wounds, and elsewhere. The sensitivity of each bacterial agent to an array of antibiotics must be tested.

The control of the environment in the intensive care unit is a major problem. Cross contamination between various patients and between patients and personnel must be constantly monitored to prevent outbreaks of infection. *See* HOSPITAL INFECTIONS.

Other acute crises. Various types of crises are managed by critical care specialists. The following are some examples. Upper gastrointestinal hemorrhage may occur because of bleeding duodenal ulcers, esophagitis, gastritis, perforations of the stomach or esophagus, or other conditions. Lower gastrointestinal bleeding may arise from conditions such as colon tumors, colon diverticula, and hemorrhoids. Therapy usually consists of rapid blood replacement and then, once the patient's condition is stable, operative control of the bleeding source. *See* HEMORRHAGE.

Acute abdominal crises. Acute catastrophic abdominal conditions can include acute pancreatitis, intestinal obstruction, perforation of a duodenal ulcer, stones of the kidney or ureter, and many other conditions. Many patients with these conditions are admitted to the intensive care unit during their preoperative diagnostic work-up and for stabilization of their cardiorespiratory and fluid and electrolyte status; they may be readmitted postoperatively for management.

Cardiac emergencies. Among the cardiac emergencies sometimes treated in intensive care units are serious abnormalities of heart rate and rhythm, toxicity from drugs used to treat heart disease, and cardiac tamponade (compression of the heart by blood or fluid in the sac surrounding it). For some conditions treatment can include emergency placement of a pacemaker.

Coma management. The sudden onset of altered mental status or coma may result from head injury, cerebrovascular accidents (stroke), drug overdose, postoperative conditions after neurosurgery, meningitis, brain tumors, and various other causes. Patients with altered mental status or coma require special nursing to prevent complications. A major consideration, especially after neurosurgical operations and head injuries, is to monitor the level of consciousness; a change in consciousness is usually the first warning of increasing intracranial pressure, which if not promptly corrected may lead to irreversible brain damage.

Acute endocrine disorders. Critical care is used in treating a number of acute endocrine catastrophies that are uncommon but potentially fatal. Examples include severe conditions caused by excessive or deficient thyroid hormone secretion, hypoparathyroidism after thyroid surgery, and adrenocortical insufficiency (Addison's disease). *See* ENDOCRINE MECHANISMS.

Injuries. Trauma from automobile accidents, penetrating injuries from violent crimes, and blunt trauma from falls or assaults pose serious technical and logistic problems that must be rapidly addressed. The injured patient must promptly receive appropriate diagnostic tests, monitoring, and therapy. Sometimes surgery must be undertaken while resuscitation efforts are still under way.

Hepatic failure and metabolic problems. Critical care also is used in managing hepatic failure and metabolic problems. Hepatic insufficiency or failure can result from conditions such as alcoholic cirrhosis, viral hepatitis, drug reactions, and poisoning. Often infection is present, as immunocompetence is impaired in liver failure. *See* CIRRHOSIS; LIVER.

Pulmonary problems. Patients whose breathing is seriously impaired by conditions such as pneumonia, asthma, pulmonary embolism (blood clot in the lung), smoke inhalation, and near-drowning may receive critical care. Measures available to assist such patients include use of mechanical ventilators (discussed above), administration of humidified supplemental oxygen through a mask or nasal prongs, and use of bronchodilator drugs. William C. Shoemaker

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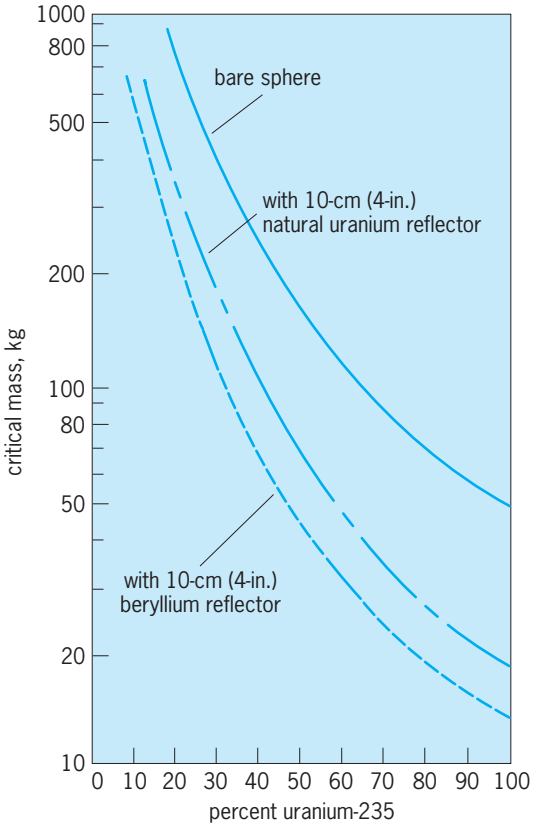
Critical mass

The amount of fissile material (uranium-233, uranium-235, or plutonium-239) that supports a self-sustaining nuclear chain reaction. It ranges from as little as 1 lb (0.45 kg) for plutonium-239 dissolved in light water to 2 tons (1800 kg) for uranium-235 fueling a 1-GWe (1 gigawatt of electric power) nuclear power reactor. *See* CHAIN REACTION (PHYSICS).

Critical mass is a definitive feature of nuclear reactors and nuclear explosives. It is increased by the presence of such neutron-absorptive materials as admixed uranium-238, aluminum pipes for flow of coolant, zirconium fuel-element cladding, and boron, cadmium, or gadolinium control rods. It is reduced by a moderator, such as graphite, heavy water, or light water, which slows down the neutrons, inhibits their escape, and thus improves the probability of producing fission in the fuel. Without a



Critical mass of a uranium sphere as a function of the uranium-235 fraction for three reflector configurations. 1 kg = 2.2 lb.

moderator, natural uranium cannot reach criticality; about 1 ton (900 kg) is used to fuel a heavy-water-moderated 1-GWe reactor. Most power reactors, especially fast reactors, are loaded with core-distributed fissile inventories that add up to many latent critical masses. Additional fissile material is also bred into the core fissile material through neutron absorption in admixed fertile material, such as in the conversion of uranium-238 to fissile plutonium-239. *See* NUCLEAR REACTOR.

Critical mass also depends on the density of the fissile component, its geometry, and the immediate surroundings. The **illustration** shows its relationship to the fissile fraction for a metallic sphere enriched in uranium-235. It also shows how a surrounding layer of suitable material (called a reflector), such as beryllium or natural uranium, can significantly reduce the critical mass. The critical mass varies for different nuclear isotopes (see **table**).

Reactor criticality. In a reactor the critical mass is not a static quantity because specific nuclear phenomena activate and influence the chain reaction. When the reactor is just critical, the reactor fission rate is stable, a condition called delayed critical. When the overall fission rate is declining or increasing, subcritical or supercritical conditions respectively exist. Subcriticality ensues if, under prevailing conditions, the fuel loading is insufficient to sustain criticality. Supercriticality (or prompt criticality) occurs at the instant that the reactive fissionable content exceeds the requirements for nuclear equilibrium.

Delayed criticality is sustained in all reactors by adjusting operating parameters that regulate the effective critical mass. Heat production for power generation, which consumes fissionable material, can be maintained by compensating with the removal of neutron absorbers through the movement of control rods, changes in the external geometry provided by reflectors, or continuous replacement of fissionable material when liquid fuel is employed. Burnable poisons and the deliberate conversion of fertile isotopes are also used to overcome reactivity losses from power-producing fuel consumption. Reactivity loss occurs also from the production of neutron-absorbing fission products, through the power-production cycle.

The criticality state of a reactor is described mathematically by the multiplication constant k, which is a ratio comparing the number of neutrons born to those consumed or lost. When the reactor is just critical, the effective value of k is unity. In calculating the value of k, all neutron sources and sinks are taken into account. The sources are (1) intrinsic yields of fission neutrons in the fissile isotope and (2) augmented fission in the fertile component. The sinks are (1) nonfission captures in the fuel, cladding, coolant, fission products, and structure and (2) leakage from the reactor.

By exploiting inherent nuclear and thermodynamic properties of core and reflector materials, reactors are designed to be almost self-regulating. The effective critical mass and state of criticality at any given moment are determined by delayed neutrons, expansion coefficients, void fractions, and resonance widths for neutron capture reactions in the reactor. Reactors are normally operated at delayed critical, with redundant features to avoid prompt criticality. Safety and power generation are dependent upon proper reactor management and conformance to operating standards that are based on these inherent design attributes. *See* DELAYED NEUTRON; REACTOR PHYSICS.

Nuclear explosions. An explosive, supercritical condition can be deliberately created by rapid introduction of reactivity in excess of the requirements for a critical mass. This is the basis for nuclear explosions, which rely on sudden explosively driven densification of a subcritical fissile component.

Approximate critical masses and sizes of fissionable spheres surrounded by 10-cm (4-in.) uranium reflectors

Isotopes	Density, g/cm^3	Critical mass, kg (lb)	Radius, cm (in.)
^{233}U	18.9	5.7 (12.6)	4.2 (1.7)
^{235}U	18.9	15.7 (34.6)	5.8 (2.3)
^{239}Pu	19.5	4.5 (9.9)	3.8 (1.5)
^{240}Pu	19.5	19.6 (43.2)	6.2 (2.4)
^{242}Pu	19.5	73 (161)	8.4 (3.3)
20% ^{235}U, 80% ^{238}U	18.9	350 (770)	9.7 (3.8)
30% ^{239}Pu + ^{241}Pu, 70% ^{240}Pu + ^{242}Pu	19.5	6 (13)	4.2 (1.7)

Tritium boosting with thermonuclear reactions is used to amplify the explosive yield of a fission chain reaction, thus reducing the fissile mass inventory of a weapon that must be compressed in order to become supercritical. *See* ATOMIC BOMB; HYDROGEN BOMB; NUCLEAR EXPLOSION.

Safety issues. Because critical mass is dependent on extrinsic factors such as geometry, moderator, and coolant flow, reactor safety issues are of extreme importance. Reactors can become prompt critical but, lacking physical mechanisms for intense compression, they are unable to explode like nuclear weapons. Early in the development of the first atomic bomb (the Manhattan Project), a few workers lost their lives in test-reactor-core prompt-criticality accidents. In 1986, at the Chernobyl reactor in the former Soviet Union, fatal consequences occurred after operators lost control of reactor criticality. The Three Mile Island reactor in Pennsylvania suffered a loss-of-coolant accident in 1979, but criticality control was maintained, no lives were lost, and only a small amount of radioactivity escaped the primary containment. *See* NUCLEAR POWER.

When subcritical quantities of nuclear materials are handled, packaged, or transported, safe geometrical configurations and environments are worked out in great detail and retained through careful management. For example, it should not be possible for a nuclear mass to become critical even if its container is flooded with water. No criticality accidents related to fuel transport have been recorded. A. DeVolpi

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Critical phenomena

The unusual physical properties displayed by substances near their critical points. The study of critical phenomena of different substances is directed toward a common theory.

Critical points. Ideally, if a certain amount of water (H_2O) is sealed inside a transparent cell and heated to a high temperature T, for instance, $T > 647$ K (374°C or 705°F), the enclosed water exists as a transparent homogeneous substance. When the cell is allowed to cool down gradually and reaches a particular temperature, namely the boiling point, the enclosed water will go through a phase transition and separate into liquid and vapor phases. The liquid phase, being more dense, will settle into the bottom half of the cell. This sequence of events takes place for water at most moderate densities. However, if the enclosed water is at a density close to 322.2 kg · m^{-3}, rather extraordinary phenomena will be observed. As the cell is cooled toward 647 K (374°C or 705°F), the originally transparent water will become increasingly turbid and milky, indicating that visible light is being strongly scattered. Upon slight additional cooling, the turbidity disappears and two clear phases, water and vapor, are found. This phenomenon is called the critical opalescence, and the water sample is said to have gone through the critical phase transition. The density, temperature, and pressure at which this transition happens determine the critical point and are called respectively the critical density ρ_c, the critical temperature T_c, and the critical pressure P_c. For water $\rho_c = 322.2$ kg · m^{-3}, $T_c = 647$ K (374°C or 705°F), and $P_c = 2.21 \times 10^7$ pascals. *See* OPALESCENCE.

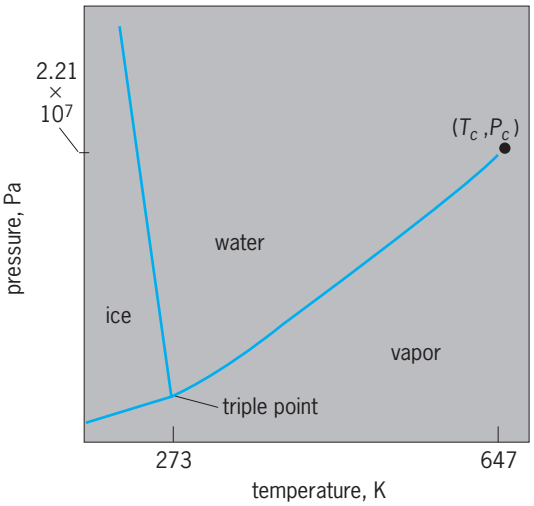

Fig. 1. Phase diagram of water (H_2O) on the pressure-temperature (*P-T*) plane.

Different fluids, as expected, have different critical points. Although the critical point is the end point of the vapor pressure curve on the pressure-temperature (*P-T*) plane (**Fig. 1**), the critical phase transition is qualitatively different from that of the ordinary boiling phenomenon that happens along the vapor pressure curve. In addition to the critical opalescence, there are other highly unusual phenomena that are manifested near the critical point; for example, both the isothermal compressibility and heat capacity diverge to infinity as the fluid approaches T_c. *See* THERMODYNAMIC PROCESSES.

Many other systems, for example, ferromagnetic materials such as iron and nickel, also have critical points. The ferromagnetic critical point is also known as the Curie point. As in the case of fluids, a number of unusual phenomena take place near the critical point of ferromagnets, including singular heat capacity and divergent magnetic susceptibility. The study of critical phenomena is directed toward describing the various anomalous and interesting types of behavior near the critical points of these diverse and different systems with a single common theory. *See* CURIE TEMPERATURE; FERROMAGNETISM.

Order parameters. One common feature of all critical phase transitions is the existence of a quantity called the order parameter. The net magnetization M is the order parameter for the ferromagnetic system. At temperatures T above T_c and under no external field, there is no net magnetization. However, as the temperature of the system is cooled slowly through T_c, a net magnetization M will appear precipitously. The increase in M is more gradual as temperature is reduced further. The nonzero magnetization is due

TABLE 1. Order parameters, theoretical models, and classification according to universality class of various physical systems*

Universality class	Theoretical model	Physical system	Order parameter
$d = 2$ $n = 1$	Ising model in two dimensions	Adsorbed films	Surface density
$n = 2$	XY model in two dimensions	Helium-4 films	Amplitude of superfluid phase
$n = 3$	Heisenberg model in two dimensions		Magnetization
$d > 2$ $n = \infty$	Spherical model	None	
$d = 3$ $n = 0$	Self-avoiding random walk	Conformation of long-chain polymers	Density of chain ends
$n = 1$	Ising model in three dimensions	Uniaxial ferromagnet	Magnetization
		Fluid near a critical point	Density difference between phases
		Mixture of liquids near consolute point	Concentration difference
		Alloy near order-disorder transition	Concentration difference
$n = 2$	XY model in three dimensions	Planar ferromagnet	Magnetization
		Helium-4 near superfluid transition	Amplitude of superfluid phase
$n = 3$	Heisenberg model in three dimensions	Isotropic ferromagnet	Magnetization
$d \leq 4$ $n = -2$		None	
$n = 32$	Quantum chromodynamics	Quarks bound in protons, neutrons, etc.	

*From K. Wilson, Problems in physics with many scales of length, *Sci. Amer.,* 241(2):158–179, August 1979.

to the partial alignment of the individual spins in the ferromagnetic substance. M is called the order parameter since the state with partial alignment of spins is more ordered than that with no alignment. The density difference in the liquid and vapor phases, $(\rho_l - \rho_v)/\rho_c$, is the proper order parameter in the fluid system; ρ_c in the denominator is the critical density. This order parameter has temperature dependence similar to that for the net magnetization M of a ferromagnetic system. For $T > T_c$ the order parameter is equal to zero, since there is only one homogeneous phase in the fluid. As the system is cooled through T_c, the fluid system phase separates with a precipitous increase in the difference of the liquid and vapor densities. A number of critical systems and their respective order parameters are listed in **Table 1**.

The order parameters assume power law behaviors at temperatures just below T_c. In the fluid and ferromagnetic systems, for $t < 0$, they follow Eqs. (1) and (2), and for $t > 0$ they obey Eqs. (3).

$$M = B(-t)^{\beta} \tag{1}$$

$$\frac{\rho_l - \rho_v}{\rho_c} = B(-t)^{\beta} \tag{2}$$

$$M = 0 \qquad \frac{\rho_l - \rho_v}{\rho_c} = 0 \tag{3}$$

In these equations, $t \equiv (T - T_c)/T_c$ reduced temperature, and β and B are respectively the critical exponent and amplitude for order parameter.

Measurement of the order parameter of fluid neon near the critical temperature [$T_c = 44.48$ K (-228.57°C or -379.43°F), $\rho_c = 484$ kg · m^{-3}, $P_c = 2.72 \times 10^6$ Pa] is shown in **Fig. 2**. The densities of the liquid (upper branch) and vapor (lower branch) phases as a function of temperature are deduced by measuring the dielectric constant of neon in the bottom and top halves of a sample cell. Data at reduced temperatures t between -4×10^{-4} and -2×10^{-3} are shown as broken lines. Careful data analysis shows that the simple power function given in Eq. (2) is not adequate to describe the data over the entire temperature range. This is the case because even in the range of reduced temperatures t between -4×10^{-4} and -4×10^{-2} the fluid is not yet inside the asymptotic or true critical region, and correction terms are needed.

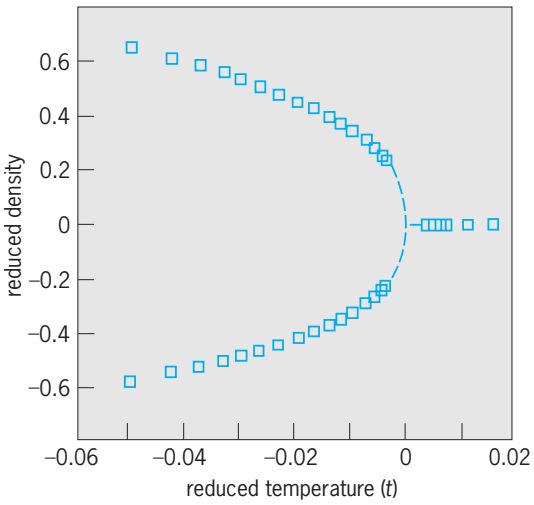

Fig. 2. Densities of liquid (upper branch) and vapor (lower branch) neon near its critical temperature. Reduced density is defined as $(\rho - \rho_c)/\rho_c$ and reduced temperature is defined as $t = (T - T_c)/T_c$. (*After M. Pestak and M. H. W. Chan, Equation of state of N_2 and Ne near their critical points: Scaling, correction to scaling, and amplitude ratios, Phys. Rev., B30:274-288, 1984*)

TABLE 2. Critical exponents and power laws in pure fluids and ferromagnets*

Thermodynamic quantity	Fluid	Ferromagnet	Power law
Specific heat	C_v	C_H	$\sim(t)^{-\alpha}$ for $t>0$ $\sim(-t)^{-\alpha'}$ for $t<0$
Order parameter = S	$(\rho_1 \pm \rho_v)\rho_c{}^{-1}$	M	= 0 for $t>0$
Response function	$K_t = \frac{1}{\rho}\left(\frac{\partial\rho}{\partial P}\right)_T$	$X = \left(\frac{\partial M}{\partial H}\right)_T$	$\sim(-t)^{\beta}$ for $t<0$ $\sim(t)^{-\lambda}$ for $t>0$ $\sim(-t)^{-\lambda'}$ for $t<0$
Critical isotherm	$P \pm P_c$	H	$\sim S\|^{\delta}\cdot$ sign of S
Correlation length	ξ	ξ	$\sim(t)^{-\nu}$ for $t>0$ $\sim(-t)^{-\nu'}$ for $t<0$
Critical correlation function of fluctuation	$G(r)$	$G(r)$	$\sim(r)^{-(d-2+\eta)}$

*d = spatial dimensionality of the critical system; H = magnetic field strength; C_v = specific heat at constant volume; C_H = specific heat at constant magnetic field strength; K_t = isothermal compressibility; X = susceptibility.

When the density data are analyzed according to the predicted functional form given by Eq. (4)

$$\frac{\rho_l - \rho_v}{\rho_c} = B(-t)^{\beta}\,[1 + B_1\,|t|^{0.5} + B_2\,|t| + \cdots] \quad (4)$$

the exponent is found to be 0.327 ± 0.002, in excellent agreement with the theoretically predicted value of 0.325 as discussed below. The value of β found for other simple fluids also converges toward the theoretical value.

Power law behavior of other quantities. The anomalous behavior of other thermodynamic quantities near the critical point can also be expressed by power laws. These can be characterized by a set of critical exponents, labeled (besides β) α, γ, δ, ν, and η: α characterizes the divergent heat capacity; γ the susceptibility (of magnets) and isothermal compressibility (of fluids); δ the critical isotherm; ν the correlation length; and η the critical correlation function. The functional forms of these power laws for the fluid and magnet systems are shown in **Table 2**. The critical opalescence phenomenon discussed above is closely related to the strong density fluctuations induced by the divergent isothermal compressibility near T_c. When the density fluctuations are correlated at lengths comparable to the wavelength of visible light, intense scattering occurs.

Mean field theories. The earliest attempts to understand the critical behavior were the van der Waals model for fluids (1873) and the Weiss model for ferromagnets (1907). These are mean field theories in the sense that the state of any particular particle in the system is assumed to be determined by the average properties of the system as a whole. In these models, all particles can be considered to contribute equally to the potential at each site. Therefore, the mean field theory essentially assumes the intermolecular interaction to be of infinite range at all temperatures. The mean field theories are qualitatively quite successful in that they predict the existence of critical points and power law dependence of the various thermodynamic quantities near the critical point. They are not quantitatively correct because the predicted values for the various exponents are not in agreement with exact model calculations or with experimental results (**Table 3**).

Scaling hypothesis. Theoretical efforts in the study of critical phenomena have been centered on predicting correctly the value of these critical exponents. One of the most important developments is the hypothesis of scaling. This hypothesis is model-independent and applicable to all critical systems. The underlying assumption is that the long-range correlation of the order parameter, such as the spin fluctuation in the ferromagnetic system and the density fluctuation in the fluid system near T_c, is responsible for all singular behavior. This assumption leads to a particular functional form for the equation of state near the critical point. With this simple assumption, it has been demonstrated that a number of inequalities among critical exponents that can be proved rigorously by thermodynamic arguments are changed into equalities. These equalities, or scaling laws, show that there are only two independent critical exponents; once two exponents in a system are given, all other exponents can be determined. What is truly impressive about this simple hypothesis is that the scaling laws, Eqs. (5)–(9), have been shown

$$\alpha = \alpha', \quad \gamma = \gamma', \quad \nu = \nu' \quad (5)$$

$$2 = \alpha + \gamma + 2\beta \quad (6)$$

$$2 = \alpha + 2\beta\delta - \gamma \quad (7)$$

$$\nu d = 2 - \alpha \quad (8)$$

$$\gamma = \nu(2 - \eta) \quad (9)$$

to be correct in almost all real and theoretical model critical systems.

The meaning of these exponents is explained above and also in Table 2; d is not an exponent but the spatial dimensionality; the primed and unprimed exponents represent the value below and

TABLE 3. Values of critical exponents

Systems	α	β	γ	δ	ν	η
Mean field model	0 (discontinuity)	$1/2$	1	3	$1/2$	0
Ising model (exact)	0 (logarithmic					
(d = 2, n = 1)	discontinuity)	$1/8$	$7/4$	15	1	0.25
Ising model (approx.)						
(d = 3, n = 1)	0.110	0.325	1.24	4.82	0.63	0.03
Heisenberg model (approx.)						
(d = 3, n = 3)	−0.10	0.36	1.38	4.80	0.705	0.03
Fluids*						
Xe	0.08 ± .02	0.325 − 0.337	1.23 ± 0.03	4.40		
SF_6		0.321 − 0.339	1.25 ± 0.03			
Ne		0.327 ± 0.002	1.25 ± 0.01			
N_2		0.327 ± 0.002	1.233 ± 0.010			
Ferromagnets* (isotropic)						
Iron, Fe	−0.09 ± .01	0.34 ± .02	1.33 ± 0.02			0.07 ± 0.07
Nickel, Ni	−0.09 ± .03	0.37 ± .03	1.34 ± 0.02	4.2 ± 0.1		

*Experimental data are the averaged measured values of a number of experiments.

above T_c, respectively. A large number of theoretical and experimental activities are concerned with putting the scaling hypothesis on a firm fundamental basis and understanding its universality and limitations.

Model systems. A great deal of insight has been gained by the construction and solution of model systems that can be solved exactly. The most famous one is the two-dimensional ($d = 2$) Ising model solved by L. Onsager in 1944. In this model, spins (little magnets) on a lattice are allowed to point in either the up or down directions, and it is assumed that only pairs of nearest-neighboring spins can interact. Onsager found a critical point for this system and calculated the values of the various critical exponents. The solution of this model is important because this is one of the very few model systems with exact solutions, and it is often used to check the validity of approximation techniques. There are many other model systems similar to the Ising model. They are distinguished from each other by the spatial (d) and spin (n) dimensionality. The spin can be oriented along an axis ($n = 1$, Ising model), or in any direction on a plane ($n = 2$, XY model), or in any direction in space ($n = 3$, Heisenberg model). *See* ISING MODEL.

These models are essentially simplified versions of real physical systems. The three-dimensional Heisenberg model ($d = 3$, $n = 3$), for example, clearly resembles the isotropic ferromagnets; the three-dimensional Ising model ($d = 3$, $n = 1$) can be found to correspond to the pure fluid system. The correspondence can be shown if the space accessible to the fluid is divided into lattice sites, and at each site the spin parameter is considered to be up if the site is occupied and down if it is not. There are other physical systems beside pure fluids that resemble the three-dimensional Ising model, for example, the binary fluid near its consolute mixing point, the uniaxial ferromagnet, and an alloy system near the order-disorder transition. A great deal of effort and ingenious mathematical techniques have been employed to obtain approximate solutions to these model systems.

Universality hypothesis. It has been observed that the measured values for the critical exponents are rather close to the calculated ones of the corresponding model system. This observation leads to the hypothesis of critical universality. According to this hypothesis, the details of the particle-particle interaction potential in the vicinity of the critical point are not important, and the critical behavior is determined entirely by the spatial dimensionality d and the spin dimensionality n. All systems, both model and real, that have the same value of d and n are said to be in the same universality class and have the same critical exponents. The hypotheses of scaling and universality are closely related: since the length of correlation between density or spin fluctuation diverges as one approaches the critical point, and any interaction potential between particles is finite in range, the details of interparticle potential are expected to be increasingly less important as one approaches the critical point. It has been shown that scaling laws can be derived from the universality hypothesis. Classification of model and physical systems according to universality classes is shown in Table 1.

Renormalization group. The renormalization group (RG) method, originally used in quantum field theory, has been introduced in the study of critical phenomena. By employing a set of symmetry transformations, the ideas contained in the principle of universality and in the scaling hypothesis can be reformulated and incorporated much more economically. As a result, a fully operational formalism is obtained from which critical exponents can be calculated explicitly. Beside the success in critical phenomena, the renormalization group method is also found to be a very useful technique in many diverse areas of theoretical physics. *See* QUANTUM FIELD THEORY.

The validity and indeed the elegance of the concept of universality and the theory of critical phenomena are borne out by experiments. As stated above, the order parameter exponents for simple fluids were found in a number of experiments to be equal to 0.327 ± 0.002. This is in excellent

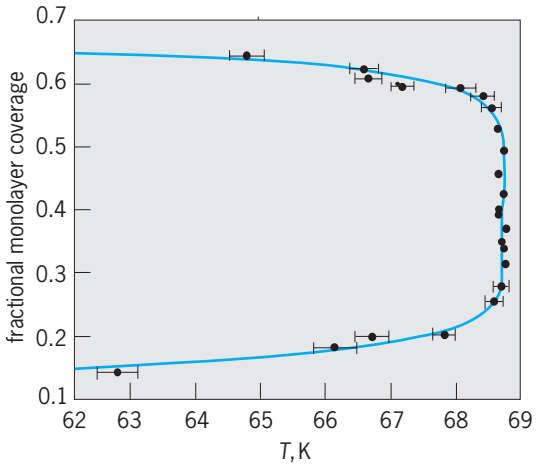

Fig. 3. Surface densities (values of fractional monolayer coverage) of monolayer methane adsorbed on graphite near its two-dimensional critical temperature. (*After H. K. Kim and M. H. W. Chan, An experimental determination of a two-dimensional liquid-vapor critical point exponent, Phys. Rev. Lett., 53:170–173, 1984*)

agreement with the value of 0.325 calculated for the three-dimensional Ising model ($n = 1$, $d = 3$) by renormalization-group and high-temperature series expansion methods. **Figure 3** shows the measured densities of the liquid and vapor phases of a two-dimensional fluid. The two-dimensional fluid is achieved by adsorbing methane molecules on an atomically flat graphite surface at monolayer and submonolayer coverages. The curve in Fig. 3, which traces the boundary separating the liquid-vapor coexistence region from the hypercritical fluid region, also reflects the liquid (upper half) and vapor (lower half) densities. It is qualitatively different from that for a bulk or three-dimensional fluid as shown in Fig. 2. When the curve is subjected to an analysis according to the power law given in Eq. (2), the resulting exponent β is found to be 0.127 ± 0.020, in excellent agreement with the value of $^1/_8$ as predicted by Onsager for the two-dimensional Ising model ($d = 2$, $n = 1$).

Dynamical effects. Besides the static critical phenomena, there are many interesting dynamical effects near the critical point, including critical slowing down, the dynamics of density and spin fluctuations, thermal and mass transport, and propagation and attenuation of sound. Understanding of these effects is far from complete.

Most critical phenomena experiments have been performed on three-dimensional systems, but a number have been done in two-dimensional and quasi-two-dimensional systems. The experiments include the order-disorder transition and the continuous melting transition of gases bound to a graphite surface and the superfluid transition of ^{4}He films on substrates. There has also been considerable interest in the influence of disorder on fluid and magnetic systems. These experiments provide interesting physical realizations of the various model systems in two dimensions. *See* PHASE TRANSITIONS; STATISTICAL MECHANICS.

Moses H. W. Chan

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Crocodile

The common name used for 15 species of large reptiles included in the family Crocodylidae, one of the three families of the order Crocodylia, which also includes the alligators, caimans, and gharials (also known as gavials). Like all crocodilians, the crocodiles are distributed throughout the tropical and subtropical regions of the world. Species occur in both saltwater and freshwater habitats. Crocodiles are generally carnivorous, feeding on invertebrates, fish, other reptiles, amphibians, birds, and mammals. A few, very narrow snouted species are believed to subsist primarily on fish. Crocodiles are primarily aquatic and nocturnal, leaving the water only to bask by day or to build their nests. Some species construct burrows into the banks of rivers or lakes where they spend part of their time. *See* ALLIGATOR; CROCODYLIA; GAVIAL; REPTILIA.

Anatomy. These animals are powerful predators with large teeth and strong jaws. Large adults of some species may exceed 20 ft (6 m) in length and are capable of overpowering and eating large grazing mammals such as deer and cattle and even, occasionally, humans. The webbed feet, laterally compressed tail, and placement of the nostrils, eyes, and ears on raised areas of the head are adaptations for an aquatic existence (see **illustration**). The raised nostrils, eyes, and ears allow the animals to float almost completely submerged while still monitoring their environment. The body, particularly the back, is covered with a series of bony plates (osteoderms) which are unfused and have no direct connection with the

American crocodile (*Crocodylus acutus*) from Ding Darling National Wildlife Refuge, Florida. (*Photo by Gerald and Buff Corsi; copyright © 2001 California Academy of Sciences*)

endoskeleton. The crocodile does not molt the epidermis in a single unit or in large pieces, as do most reptiles.

Size. There is considerable controversy over just how large crocodiles may grow. The saltwater crocodile (*Crocodylus porosus*) of the Indo-Pacific and Australia; the Nile crocodile (*C. niloticus*) of Africa; the American crocodile (*C. acutus*; see illustration) of Central and South America, many Caribbean islands, and Florida; and the Orinoco crocodile (*C. intermedius*) of the Orinoco drainage in South America are reliably reported to approach or slightly exceed 20 ft (6 m) in total length. Johnson's crocodile (*C. johnsoni*) of Australia; the Mugger crocodile (*C. palustris*) of India and Sri Lanka; the Siamese crocodile (*C. siamensis*) of Malaysia and Indonesia; the New Guinea and Philippine crocodiles (*C. novaeguineae* and *C. novaeguinea mindorensis*) of Papua New Guinea and the Philippines; Morelet's crocodile (*C. moreletti*) of Mexico, Guatemala, and Belize (British Honduras); the Cuban crocodile (*C. rhombifer*) of Cuba; and the narrow-snouted crocodile (*C. cataphractus*) of Africa have a maximum length of between 9 and 14 ft (3 and 4 m). The false gavial (*Tomistoma schlegeli*) of Malaysia, Borneo, and Sumatra may occasionally reach 16 ft (5 m) in length. The broad-snouted crocodiles (*Osteolaemus tetraspis*) of west-central Africa seldom exceed 6 ft (2 m) in total length.

Reproduction. Reproduction in crocodiles, and in all crocodilians, is elaborate and has been well studied in a few species. Courtship and mating occur in the water. The female digs a hole in the soil for the 10 to 50 eggs, or, in the case of several species, constructs an elevated nest of available soil and vegetation near the water. The decaying vegetation gives off heat that aids in the incubation of the eggs. Nests may be as large as 12 ft (3.5 m) across and 4 ft (1.2 m) high and are often guarded by the female, which may also construct wallows adjacent to the nest where she remains for much or all of the incubation period. Incubation takes approximately 2 months. In some species, at least, the hatchlings croak or grunt from within the nest, and the female opens the nest to assist their emergence. The young may remain together as a pod with the female for a year or more. Hearing and vocal communication are well developed in the crocodiles, and a variety of bellows, snarls, and grunts are utilized in their elaborate social behavior. *See* REPRODUCTIVE SYSTEM.

Conservation and ecology. Many factors have contributed to the overall decline of crocodiles worldwide. Crocodile hides are considered valuable, and the decline of many species may be attributed to hunting and poaching. The loss of habitat through land development for other uses also contributes significantly to the decline of crocodile species. Most countries with native crocodile populations are implementing conservation measures, and international efforts are being made to regulate trade in crocodile products. Papua New Guinea has developed a conservation-farming program for its species which, if successful, could serve as a model for the conservation of crocodiles and all crocodilians throughout the world.

The Siamese crocodile (*C. siamensis*), found in Thailand, is considered to be the most endangered crocodile in the world and is listed as "critically endangered" by the International Union for the Conservation of Nature and Natural Resources (IUCN). In the United States, the American alligator (*Alligator mississippiensis*) and the American crocodile (*C. acutus*; see illustration) are both considered to be in decline. The American alligator is listed as "federally threatened" in certain parts of its range, whereas the American crocodile, found in the coastal contact zone of southern Florida, is listed as "critically endangered." Conservation efforts are hindered by the lack of knowledge about the population biology of crocodiles. Historically, they existed in many areas in large numbers and were evidently of considerable importance in the energetics and nutrient cycling in the habitats where they occurred. The overall decline of some species makes studies of their ecological importance difficult.

Howard W. Campbell; W. Ben Cash

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Crocodylia

An order of the class Reptilia (infraclass Archosauria) which is aquatic species including the alligators, caimans, crocodiles, and gavials. The group has a long fossil history from the Late Triassic, and its members are the closest living relatives of the extinct dinosaurs and the birds. The 23 living species are found in tropic areas of Africa, Asia, Austria, and the Americas. One form, the saltwater crocodile (*Crocodylus porosus*), has traversed oceanic barriers from the East Indies as far east as the Fiji Islands. *See* ARCHOSAURIA; REPTILIA.

Morphology. The order is distinguished from other living reptiles in that it has two temporal foramina, an immovable quadrate, a bony secondary palate, no shell, a single median penis in males, socketed teeth, a four-chambered heart, and an oblique septum that completely separates the lung cavities from the peritoneal region.

Certain of these unique features and other salient characteristics of the Crocodylia are intimately associated with their aquatic life. The secondary palate, composed of medial expansions of premaxillary, maxillary, palatine, and pterygoid bones, divides the mouth cavity into two separate regions. The area above the bony palate forms an air passage extending from the external nostrils at the tip of the snout

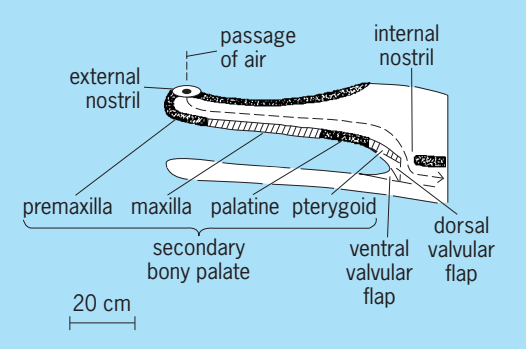

Fig. 1. Sagittal section of crocodylian head to show secondary palate and air passage.

posteriorly to near the orbital region. The lower space retains the usual functions of the mouth and is bordered above by the secondary palate, and below by the lower jaw. A special pair of fleshy flaps are found at the posterior end of the mouth cavity and form a valvular mechanism which separates the mouth from the region where the air passage opens into the throat (**Fig. 1**). This complex arrangement allows crocodylians to breathe even though most of the head is under water, or the mouth is open holding prey or full of water. Inasmuch as these great reptiles are extremely active when submerged and may remain under water for considerable periods, the increased efficiency of respiration provided by the oblique septum is an obvious adaptation. The significance of the four-chambered heart also lies in its contribution to rapid cellular respiration through an increased efficiency of circulation. Interestingly, all three of these structural characteristics are also found in mammals, although differing in detail. The Crocodylia, however, are not closely related to any of the reptiles that ultimately gave rise to the mammal stock.

Other aquatic adaptations include valvular external nostrils that are closed during immersion, a recessed eardrum that can be covered by a skin flap underwater, webbed toes, and a long, compressed, muscular tail that propels the crocodylian through the water with strong lateral thrusts. In addition, the eyes, adapted for nocturnal activity by a vertically elliptical pupil, and the nostrils are mounted in raised areas of the head so that the animal may see and breathe without exposing much of its body above water. The method of locomotion on land is by progression on all fours, with the belly and head held off the ground and the tail dragging behind. In the water, movement is produced by lateral undulations of the tail and the forelimbs are held flat against the sides of the body.

Feeding. Crocodylians are carnivorous and prey upon insects, mollusks, fishes, and mammals. Usually only the larger adults attack terrestrial prey, which is usually caught at the water's edge and after being grasped in the powerful jaws is dragged down to be drowned. The jaws, numerous sharp teeth, and the habit of twisting and rolling underwater after the prey is seized all contribute to the particular reputation crocodilians share. Even on land they are imposing creatures, especially when utilizing the muscular tail for defense. As with any large predator, crocodilians should be treated with respect and caution should be used in approaching individuals in the wild.

Reproduction. During the breeding season male crocodilians set up territories on land which they defend against intruders of the same species. During this period their loud roars are frequently heard at night. Fertilization is internal and the shelled eggs are deposited in excavations in the sand or in large nests of decaying vegetation, depending upon the species. In some forms the female guards the nest and several females may take turns in protecting a communal nest. It has been reported that when the young hatch the female may liberate them from the nest and lead them to the nearest water.

Classification. The living species of crocodilians are placed in two families and eight genera. The family Crocodylidae contains two subgroups: the true crocodiles, Crocodylinae, including the genera *Crocodylus* found in all tropic areas, *Osteolaemus* in central Africa, and the false gavial (*Tomistoma*) in Malaya and the East Indies; the alligators and caimans, Alligatorinae, including the genera

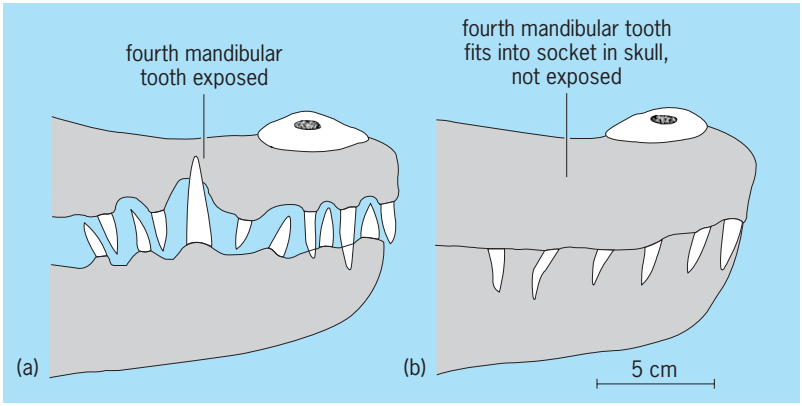

Fig. 2. Comparative lateral view of crocodylian snouts, showing tooth arrangement of (*a*) Gavialidae and Crocodylinae, and (*b*) Alligatorinae.

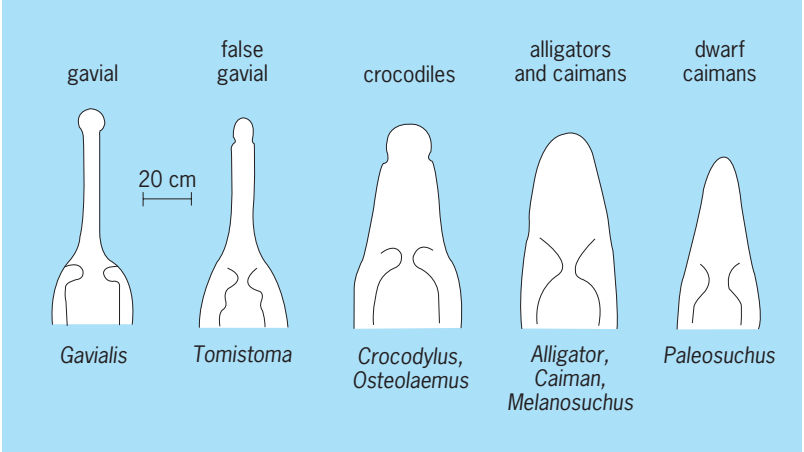

Fig. 3. Dorsal view of crocodylian heads.

Alligator of the southeastern United States and near Shanghai, China, the *Caiman* from Central and South America, and *Melanosuchus* and *Paleosuchus* of South America. The gavial (*Gavialis gangeticus*) of India and north Burma is the only living member of the family Gavialidae. Crocodiles differ most obviously from alligators and caimans in head shape and in the position of the teeth, although other technical details also separate them (**Figs. 2** and **3**). The gavial differs from other living forms in its extremely long and narrow snout, with 27 or more teeth on the upper jaw and 24 or more on the lower. Even the false gavial (*Tomistoma*), which resembles the gavial in head shape, has no more than 21 teeth on the upper jaw and 20 on the lower. *See* ALLIGATOR; CROCODILE; GAVIAL.

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Crocoite

A mineral with the chemical composition $PbCrO_4$. Crocoite occurs in yellow to orange or hyacinth red, monoclinic, prismatic crystals with adamantine to vitreous luster (see **illus.**); it is also massive granular. Fracture is conchoidal to uneven. Hardness is 2.5–3 on Mohs scale and specific gravity is 6.0. Streak, or color of the mineral powder, is orangish-yellow. It fuses easily.

Needlelike crocoite crystals in association with cerussite and dundasite, Dundas, Tasmania. (*Specimen courtesy of Department of Geology, Bryn Mawr College*)

Crocoite is a secondary mineral associated with other secondary minerals of lead such as pyromorphite and of zinc such as cerussite. It has been found in mines in California and Colorado. *See* LEAD.

Edward C. T. Chao

Crossing-over (genetics)

The process whereby one or more gene alleles present in one chromosome may be exchanged with their alternative alleles on a homologous chromosome to produce a recombinant (crossover) chromosome which contains a combination of the alleles originally present on the two parental chromosomes. Genes which occur on the same chromosome are said to be linked, and together they are said to compose a linkage group. In eukaryotes, crossing-over may occur during both meiosis and mitosis, but the frequency of meiotic crossing-over is much higher. This article is concerned primarily with meiotic crossing-over. *See* ALLELE; CHROMOSOME; GENE; LINKAGE (GENETICS).

Crossing-over is a reciprocal recombination event which involves breakage and exchange between two nonsister chromatids of the four homologous chromatids present at prophase I of meiosis; that is, crossing-over occurs after the replication of chromosomes which has occurred in premeiotic interphase. The result is that half of the meiotic products will be recombinants, and half will have the parental gene combinations (**Fig. 1**). Using maize chromosomes which carried both cytological and genetical markers, H. Creighton and B. McClintock showed in 1931 that genetic crossing-over between linked genes was accompanied by exchange of microscopically visible chromosome markers. *See* RECOMBINATION (GENETICS).

During meiosis, crossing-over occurs at the pachytene stage, when homologous chromosomes are completely paired. At diplotene, when homologs separate, the sites of crossing-over become visible as chiasmata, which hold the two homologs of a bivalent together until segregation at anaphase I. Each metaphase I bivalent will necessarily have at least one chiasma. In favorable material, such as grasshopper spermatocytes, it is possible to observe that each diplotene chiasma involves a crossover of two of the four chromatids at one site.

Where two or more crossovers occur in one bivalent, they usually do not cluster together but are widely separated; this is known as chiasma interference. The occurrence of one crossover event appears to preclude the occurrence of a second crossover in the immediate vicinity. In addition, the distribution of occurrence of chiasmata along a chromosome may be localized; the probability that a crossover will occur is higher in some chromosome segments and lower in other segments.

In general, the closer two genes are on a chromosome, that is, the more closely linked they are, the less likely it is that crossing-over will occur between them. Thus, the frequency of crossing-over between different genes on a chromosome can be used to produce an estimate of their order and distances apart; this is known as a linkage map. *See* GENETIC MAPPING.

Molecular mechanisms. Since each chromatid is composed of a single deoxyribonucleic acid (DNA) duplex, the process of crossing-over involves the

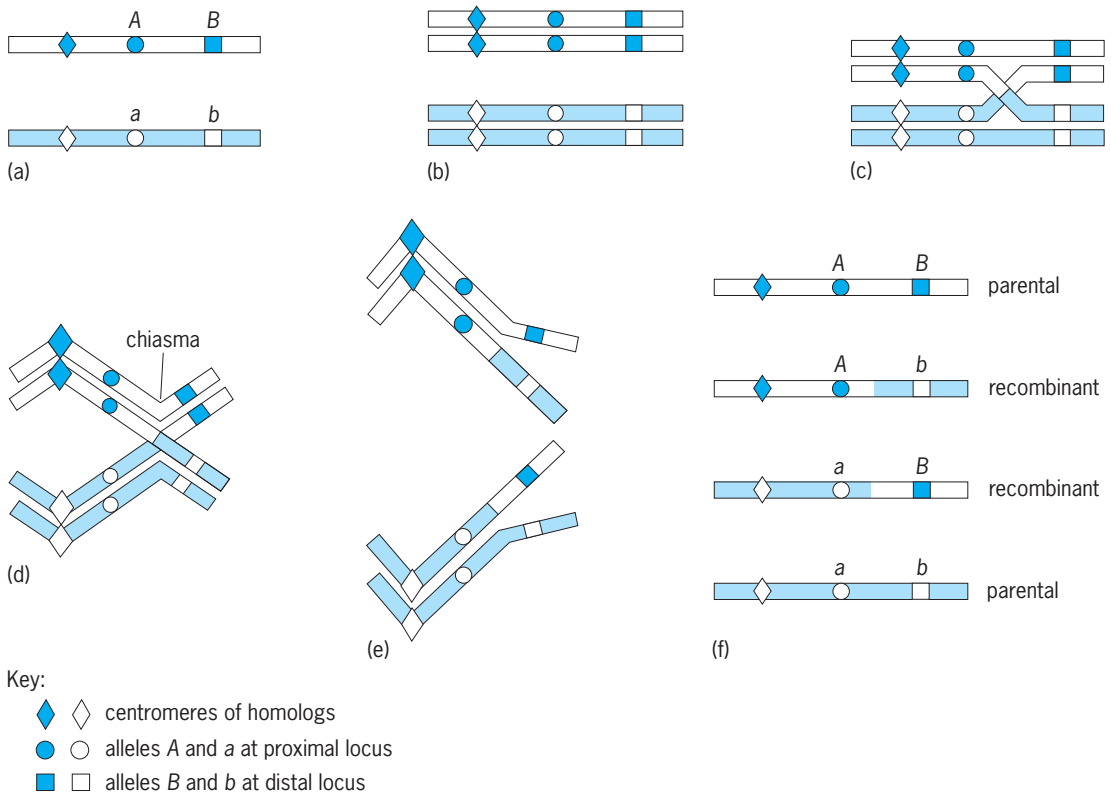

Fig. 1. Meiotic crossing-over between two gene loci *A* and *B* on homologous chromosomes with different alleles at each locus. (*a*) Prior to replication. (*b*) After replication in premeiotic interphase. (*c*) Crossing-over at pachytene between nonsister chromatids of paired homologs (bivalent). (*d*) Diplotene bivalent held together by a chiasma at crossover point. (*e*) Anaphase I segregation of chromosomes in bivalent. (*f*) Four meiotic products after anaphase II segregation. In general, centromeres and loci proximal to the chiasma (crossover) segregate at first division, while loci distal to the chiasma segregate at second division.

breakage and rejoining of DNA molecules. Although the precise molecular mechanisms have not been determined, it is generally agreed that the following events are necessary: (1) breaking (nicking) of one of the two strands of one or both nonsister DNA molecules; (2) heteroduplex (hybrid DNA) formation between single strands from the nonsister DNA molecules; (3) formation of a half chiasma, which is resolved by more single-strand breakages to result in either a reciprocal crossover, a noncrossover, or a nonreciprocal crossover (conversion event).

Two molecular models of recombination which have gained credence are those of R. Holliday and of M. Meselson and C. Radding. Holliday's model postulates nicks in both chromatids at the initiation of crossing-over (**Fig. 2**). Meselson and Radding postulate single-strand cut in only one DNA strand. Repair synthesis displaces this strand, which pairs with its complement on the other chromatid, thereby displacing and breaking the other strand of that DNA molecule. Following pairing and ligation of the two remaining broken ends, a half chiasma is formed. Other models have been postulated in which recombination is initiated by a double-stranded break in one chromatid. In all the above models, gene conversion can occur in the middle region of the molecules (with or without outside marker crossing-over) by mismatch repair of heteroduplex DNA.

Ultrastructural cytology. Pachytene, the meiotic stage at which crossing-over is considered to occur, corresponds with the period of close pairing or synapsis of homologous chromosomes. Electron microscopy has revealed that proteinaceous structures, the synaptonemal complexes (**Fig. 3**), are involved in the synapsis of chromosomes. A synaptonemal complex forms during zygotene by pairing of axial elements from two homologous chromosomes. It is present along the whole length of each pachytene bivalent and disappears at diplotene. Evidence from inhibitor studies and mutant stocks shows that the synaptonemal complex is necessary for meiotic crossing-over to occur. However, in cases such as desynaptic mutants, some hybrids, and the female silkworm, complete pachytene synaptonemal complexes have been observed, but no crossing-over occurs, showing that the synaptonemal complex alone is not sufficient to cause crossing-over.

In *Drosophila melanogaster* oocytes, the occurrence at pachytene of dense spherical bodies bridging the central region of the synaptonemal complex has been described. These bodies coincided in number and position with expected crossover events, and therefore were named recombination nodules. A variety of oval and bar-shaped recombination nodules (Fig. 3) have also been found in organisms as diverse as fungi, humans, rat, silkworm, and maize. In many cases their number correlates with crossover

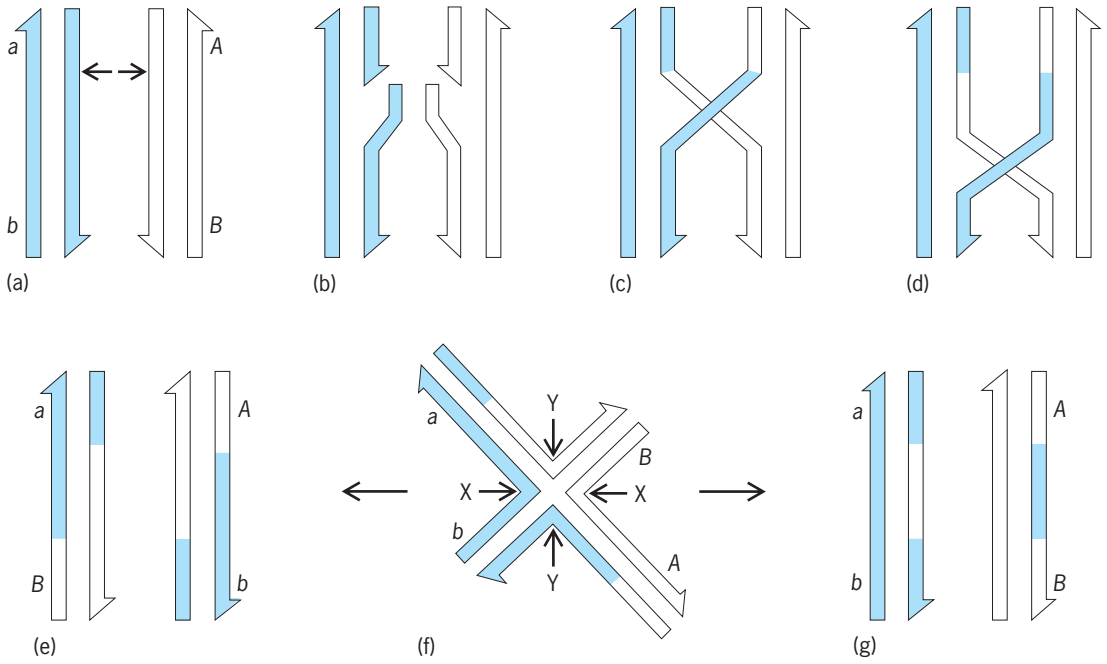

Fig. 2. Molecular model of recombination, based on that of R. Holliday. Only the two recombinant chromatids are shown. (*a*) Two homologous nonsister chromatid DNA molecules, with gene loci *ab* and *AB*, respectively; arrows are paired nicking sites. (*b*) After single-strand nicks at equivalent sites in both chromatids, nicked strands separate. (*c*) Nicked strands are displaced and reanneal to opposite duplex forming a half chiasma. (*d*) Migration of half chiasma increases length of heteroduplex DNA. (*e*) Isomerization of structure in *d* by rotation gives open half-chiasma form; paired nicks occur in two strands. (*f*) Nicks at X,X resolve the half chiasma into reciprocal crossover chromatids *aB* and *Ab*, with heteroduplexes in middle regions; or (*g*) nicks at Y,Y resolve the half chiasma into noncrossover chromatids *ab* and *AB*, with heteroduplexes in middle regions.

frequency. It has been suggested that recombination nodules are prerequisites for crossing-over. If this is so, the recombination nodule may represent a complex of enzymes involved in the early events of recombination (nicking, strand separation, repair synthesis).

DNA repair synthesis has been observed during pachytene in lily microsporocytes, and has been shown to be reduced in an achiasmatic mutant. Prophase I of lilies is characterized by the presence of several proteins which could have a role in crossing-over, for example, DNA binding protein, endonucleases, ligases, and kinase. Inhibition of protein synthesis at zygotene-pachytene results in failure of crossing-over. Thus both DNA synthesis and protein synthesis appear necessary for meiotic crossing-over in lily. *See* MEIOSIS; MITOSIS.

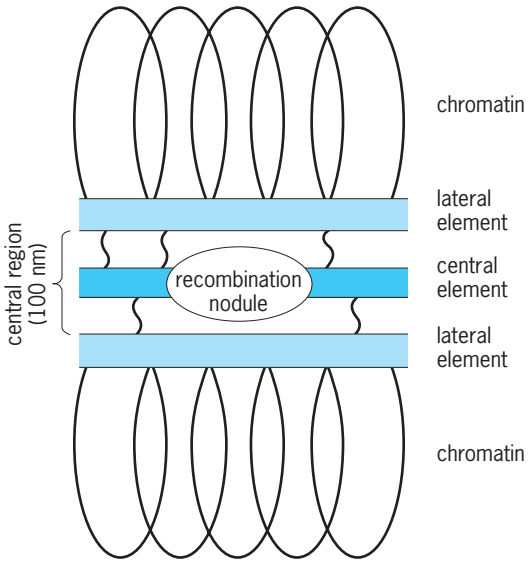

Fig. 3. Longitudinal section of a synaptonemal complex with a recombination nodule.

Sex chromosomes. The differentiated X and Y sex chromosomes in human males and many animals (Z and W chromosomes in female birds) have small regions near one tip which undergo pairing and crossing-over at meiotic prophase I. Electron microscopy of the pachytene XY reveals the formation of a short synaptonemal complex segment with a recombination nodule in the majority of cases; the presence of a chiasma between the X and Y at metaphase I indicates the occurrence of crossing-over. An obligatory crossover in the XY bivalent is necessary to ensure regular segregation of X and Y to opposite poles at anaphase I. The pairing region contains a few gene loci on both X and Y chromosomes which exhibit an autosomallike inheritance pattern. Recombination between genes and DNA sequences in this pseudoautosomal region confirms the occurrence of obligatory crossing-over. The rare occurrence of XX males in some cases is accounted for by abnormal recombination events outside the pseudoautosomal region which have transferred the male sex-determining gene from the Y to the X chromosome. *See* SEX DETERMINATION; SEX-LINKED INHERITANCE.

C. B. Gillies

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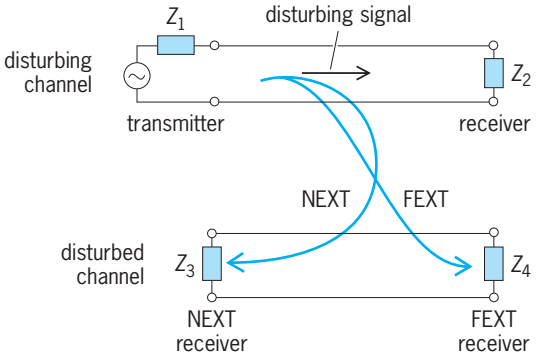
Crosstalk

Interference in a communications channel (disturbed channel) caused by activity in other communications channels (disturbing channels). The term was originally used to denote the presence in a telephone receiver of unwanted speech signals from other conversations, but its scope has been extended by common usage to include like effects in other types of communications.

Causes. The cause of crosstalk is some form of coupling mechanism between the disturbed channel and the disturbing channels. Communications channels are normally segregated by space, frequency, time, code, or polarization division, or by some combination of the five, to avoid such coupling, but economic and other constraints often preclude complete segregation.

Space division. In space-division segregation, each communication channel is assigned its own transmission medium, for example, a pair of wires is a multipair cable (see **illus.**) or a separate radio propagation path. Coupling between channels is caused by the physical proximity and relative orientation of the transmission media. The coupling is usually electromagnetic, a linear phenomenon that is independent of signal level in the channels. *See* ELECTROMAGNETIC COMPATIBILITY.

Frequency division. In frequency-division segregation, a single, common transmission medium (for example, a coaxial cable system) is used to provide many communications channels, and each channel is assigned a separate frequency band within the medium. A signal is placed in its channel by amplitude modulation (AM) or angle modulation frequency modulation (FM) or phase modulation (PM). With AM systems, nonlinearities in the transfer characteristics of system elements create harmonic, or intermodulation, distortion. Typically, nonlinearities are introduced by system amplifiers that are used to compensate for loss in the medium. Intermodulation products (distortions) have frequencies that are combinations (sums, differences, multiples) of the frequencies of the signals being carried by the system. Thus, nonlinearities spawn many coupling paths among the channels of the system. The coupling is nonlinear, that is, the losses of the paths are functions of signal amplitudes. In angle-modulation systems, the principal cause of intermodulation distortion is frequency dependency in the amplitude and delay characteristics of the transmission medium (including amplifiers), but variation of the delay characteristic of the medium as a function of signal level (AM to PM conversion) can also cause intermodulation distortion. The resulting coupling paths among the channels of the system are similar to those in an AM system. *See* AMPLITUDE MODULATION; DISTORTION (ELECTRONIC CIRCUITS); FREQUENCY MODULATION; MODULATION; PHASE MODULATION.



Crosstalk in a space-division system. Z_1, Z_2, Z_3, and Z_4 designate impedances.

Time division. In time-division segregation, a single common transmission medium is time-shared to provide many communications channels, with each channel allotted a separate time interval. These may be assigned on the basis of relative activity among the channels, or each channel may be periodically assigned a like interval regardless of activity, which is usually the case with digital systems. Interchannel coupling is the result of the presence of vestigial energy in the transmission medium from previous channel assignments. The coupling tends to be linear (the loss tends to be independent of signal amplitude), and often involves an impedance that is common to both the disturbing and disturbed channels. *See* PULSE MODULATION.

Code division. In code-division segregation, which is generally used by spread-spectrum communications systems, each communications channel is modulated by a coding sequence that is orthogonal to the coding sequences of the other channels. Perfect orthogonality is generally not achieved due to imperfections, like finite-length coding sequences, which cause some crosstalk. *See* SPREAD SPECTRUM COMMUNICATION.

Polarization division. In polarization-division segregation, transmissions are isolated from each other by use of opposite polarizations (for example, right-hand versus left-hand circular polarizations or vertical versus horizontal linear polarizations). Perfect cross-polarization isolation is not attainable due to antenna effects, antenna alignments, and variations of the transmission medium, leading to some degree of crosstalk. *See* POLARIZATION OF WAVES.

Classification. Crosstalk is classified in a variety of ways. The type of coupling (electromagnetic, intermodulation, or common impedance) indicates the

mechanism. The terms near-end crosstalk (NEXT) and far-end crosstalk (FEXT) indicate the relative directions of signal propagation in the disturbed and disturbing channels (see illus.). The terms direct crosstalk, where the disturbing channel couples directly to the disturbed channel, and indirect crosstalk, where the coupling path between disturbing and disturbed channels involves a third, or tertiary, channel, are often used to further describe crosstalk caused by electromagnetic coupling. Interaction crosstalk (IXT) is a term used to further describe indirect crosstalk that couples from the disturbing channel to the tertiary channel at one place, propagates along the tertiary channel, and subsequently couples into the disturbed channel at another place. Transverse crosstalk is a term that includes all direct and indirect crosstalk that is not interaction crosstalk.

Intelligible crosstalk is understood by the receiver of the disturbed channel, whereas nonintelligible crosstalk is not. Intelligible crosstalk most often occurs between channels carrying similar analog signals, such as analog speech signals. Intelligible crosstalk between channels, one of which is carrying a digital signal, or between channels carrying different types of analog signals (a voice signal and a video signal) is less likely to occur because the receiver does not comprehend the information in the disturbing signal. Nonintelligible crosstalk often affects the receiver in much the same way as noise, but in some cases it may be nearly intelligible, for example, cross-talk from one analog speech channel to another where intermodulation coupling has inverted the frequency spectrum without masking the syllabic nature of the disturbing speech signal. *See* ELECTRICAL NOISE.

Quantitative measures. The magnitude of crosstalk that becomes disruptive to a communications channel depends on many factors: the type of signals involved, the intelligibility of the crosstalk, the acuity of the receiver, the activity factors for the channels, and so on. In some cases crosstalk affects the disturbed channel in much the same way as noise, and signal-to-noise ratios are applicable measures. Digital channels tend to be of this nature. Typically, binary (two-level) digital signals require peak signal–to–root-mean-square noise ratios equal to, or better than, about 20 dB (voltage or current ratios of 10:1 or better). Crosstalk from one analog video channel to another produces an effect much different from noise, but a signal-to-interference ratio is an appropriate measure because the primary signal is usually present. Peak signal-to-peak interference ratios of 60 dB or better (voltage or current ratios of 1000:1 or better) are generally required for good reception of such signals. *See* SIGNAL-TO-NOISE RATIO.

In other cases, such as intelligible crosstalk in telephony, signal-to-crosstalk ratios are not applicable measures because the primary signal is not always present to mask the interference. In telephony, detection of intelligible crosstalk by the receiver is particularly disruptive—it does not really distort the primary signal, but it represents a breach of privacy—the detection of nonintelligible crosstalk that seems to be intelligible (of a syllabic nature) is almost as disruptive. Ever-present noise in the disturbed channel tends to mask crosstalk detection. To prevent detection, the noise power must exceed the crosstalk power by about 10 dB (noise-to-crosstalk power ratio of 10:1). When a disturbed telephony channel has very little noise, the typical receiver can detect crosstalk levels of about −85 dBm (−85 decibels below 1 milliwatt, or 3×10^{-9} mW). For comparison, a typical level for the primary signal at the receiver is about −25 dBm (3×10^{-3} mW). Objectives for the crosstalk performance of a telephony network are often expressed by crosstalk index, a measure of the probability of the detection of intelligible crosstalk during the course of a telephone call. Values for crosstalk index are expressed in percentages, and typical values are less than 1%. *See* DECIBEL.

Remedies. Most remedies for reducing crosstalk entail some technique for decreasing coupling among the communications channels involved. The use of twisted pairs in multipair cables, of shielding for each pair, or of coaxial conductors and optical fibers in place of pairs are common techniques for reducing electromagnetic coupling where space division alone is not adequate. Improved control of signal levels and improved linearity in amplifiers are effective for frequency-division systems. Often such improvements are made possible by advances in technology. Selection of the appropriate type of modulation (AM, FM, PM) is also important. Crosstalk within multichannel digital systems that transport digital versions of analog signals can be reduced with the help of a separate coder-decoder for each analog channel in place of a common, time-shared coder-decoder for all channels. *See* ELECTRICAL SHIELDING; OPTICAL FIBERS.

Signal processing in the disturbed channel can sometimes be effective in reducing crosstalk. Syllabic compandors, which are sometimes used with telephony channels on analog carrier systems to reduce noise in silent intervals during conversations, are effective in reducing crosstalk as well. Pulse shaping (modification of the signal spectrum) can be effective in reducing crosstalk between digital transmission systems and in reducing intersymbol interference within a digital transmission system. One of the most important tools for control of crosstalk is spectrum management, a systematic approach to the assignment of frequency bands, signal levels, and the like in such way as to offer the most efficient use of the transmission media. *See* ELECTRICAL COMMUNICATIONS; ELECTRICAL INTERFERENCE; RADIO SPECTRUM ALLOCATIONS; TELEPHONE SERVICE.

Jonathan W. Smith

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Crown gall

A neoplastic disease of primarily woody plants, although the disease can be reproduced in species representing more than 90 plant families. The disease results from infection of wounds by the free-living soil bacterium *Agrobacterium tumefaciens* which is commonly associated with the roots of plants.

The first step in the infection process is the site-specific attachment of the bacteria to the plant host. Up to half of the bacteria become attached to host cells after 2 h. At 1 or 2 weeks after infection, swellings and overgrowths take place in tissue surrounding the site of infection, and with time these tissues proliferate into large tumors (see **illus.**). If infection takes place around the main stem or trunk of woody hosts, continued tumor proliferation will cause girdling and may eventually kill the host. Crown gall is therefore economically important, particularly in nurseries where plant material for commercial use is propagated and disseminated.

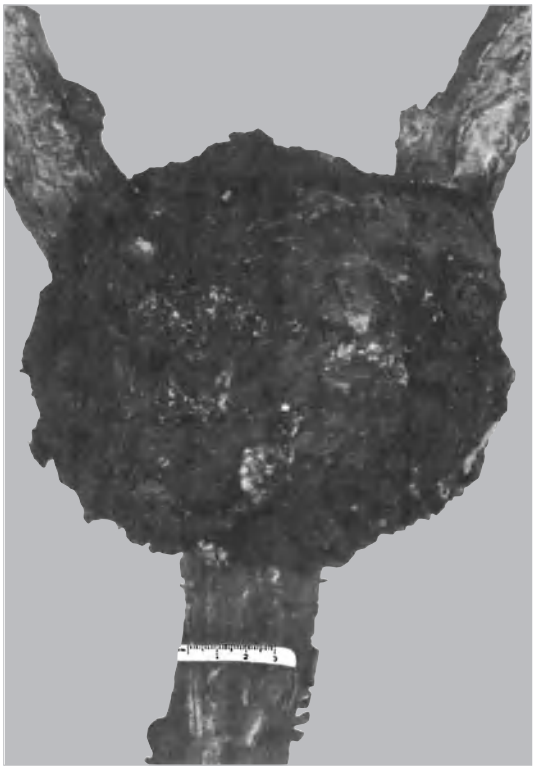
Crown gall tumors can be cultured free of bacteria on synthetic, defined media and, under proper conditions, can reinitiate tumors when grafted back onto healthy plants. Unlike healthy normal cells, crown gall tumor cells do not require an exogenous source of phytohormones (auxins and cytokinin) for growth in culture because they readily synthesize more than sufficient quantities for their own growth. The tumor cells also synthesize basic amino acids, each conjugated with an organic acid, called opines. Octopine and nopaline are opines that are condensation products of arginine and pyruvic acid, and arginine and 2-ketoglutaric acid, respectively. The tumor cells also grow about four times faster and are more permeable to metabolities than normal cells.

These cellular alterations, such as the synthesis of opines and phytohormone regulation, result from bacterial genes introduced into host plant cells by *A. tumefaciens* during infection. Although it is not understood how these genes are introduced into the plant cell, the genes for the utilization of these opines and for regulating phytohormone production have been found to be situated on an extrachromosomal element called the pTi plasmid. This plasmid, harbored in all tumor-causing *Agrobacterium* species, also carries the necessary genetic information for conferring the tumor-inducing and host-recognition properties of the bacterium. It is thus possible to convert avirulent strains of *A. tumefaciens* and related species to virulent forms by introducing the pTi plasmid into them through standard microbial genetic procedures. Also, the host range of a given *Agrobacterium* can be modified by introducing a pTi plasmid from a bacterial strain having a different host range.

Analyses made of crown gall nuclear genetic material (DNA) have revealed the presence of a specific segment of pTi plasmid DNA known as the T-DNA. Several distinct genes such as chloramphenicol acetyltransferase, and β-galactosidase have been inserted into the T-DNA by recombinant DNA methods. These genes can be introduced into plants through infection by *A. tumefaciens* carrying a pTi plasmid with one of these genes inserted in an appropriate location within the T-DNA. These molecular and genetic manipulations have made it feasible to genetically modify plants for desirable traits, a procedure popularly called genetic engineering.

Transfer of the T-DNA requires a set of genes that are clustered in another region of the pTi plasmid, known as the *vir* region. *Vir* region genes are essential for virulence of *A. tumefaciens*, and seem to respond to signals from wounded plant tissue. Although these signals remain to be identified, the interactions with the *vir* genes suggest a long-time association and dependency of *A. tumefaciens* with its host plants. Consequently, crown gall is a result of this unique bacteria-plant interaction, whereby *A. tumefaciens* genetically engineers its host to produce undifferentiated growth in the form of a large tumor, in which there is the synthesis of a unique food source in the form of an opine for specific use by the bacterial pathogen. *See* BACTERIAL GENETICS; GENETIC ENGINEERING; PLANT HORMONES; PLANT PATHOLOGY.

Clarence J. Kado



Crown gall on peach.

Crushing and pulverizing

The reduction of materials such as stone, coal, or slag to a suitable size for their intended uses such as road building, concrete aggregate, or furnace firing. Crushing and pulverizing are processes in ore dressing needed to reduce valuable ores to the fine size at which the valueless gangue can be separated

Crushing specifications

Category	Feed	Product	Reduction ratio	Equipment used
Primary crushing	27–12 in. (69–30 cm)	9–4 in. (23–10 cm)	3:1	Jaw, gyratory, cone
Secondary crushing (one or two stages)	9–4 in. (23–10 cm)	1–1/2 in. (25–13 mm)	9:1	Hammer mill, jaw, gyratory, cone, smooth roll, and toothed rolls
Pulverizing	1–1/2 in. (25–13 mm)	60–325 mesh	60:1	Ball and tube, rod, hammer, attrition, ball race, and roller mills

from the ore. These processes are also used to reduce cement rock to the fine powder required for burning, to reduce cement clinker to the very fine size of portland cement, to reduce coal to the size suitable for burning in pulverized form, and to prepare bulk materials for handling in many processes. *See* MATERIALS-HANDLING EQUIPMENT; ORE DRESSING.

Equipment suitable for crushing large lumps as they come from the quarry or mine cannot be used to pulverize to fine powder, so the operation is carried on in three or more stages called primary crushing, secondary crushing, and pulverizing (see **table**). The three stages are characterized by the size of the feed material, the size of the output product, and the resulting reduction ratio of the material. The crushing-stage output may be screened for greater uniformity of product size.

Reduction in size is accomplished by five principal methods: (1) crushing, a slow application of a large force; (2) impact, a rapid hard blow as by a hammer; (3) attrition, a rubbing or abrasion; (4) sudden release of internal pressure; and (5) ultrasonic forces. The last two methods are not in common use.

Crushers. All the crushers used in primary and secondary crushing operate by crushing as defined above except the hammer mill, which is largely impact with some attrition.

Primary crushers. The Blake jaw crusher using a double toggle to move the swinging jaw is built in a variety of sizes from laboratory units to large sizes having a feed inlet 84 by 120 in. (213 by 305 cm; **Fig. 1**). Large units have a capacity of over 1000 tons/h (900 metric tons/h) and produce a 9-in. (23-cm) product. The Dodge jaw crusher uses a single toggle

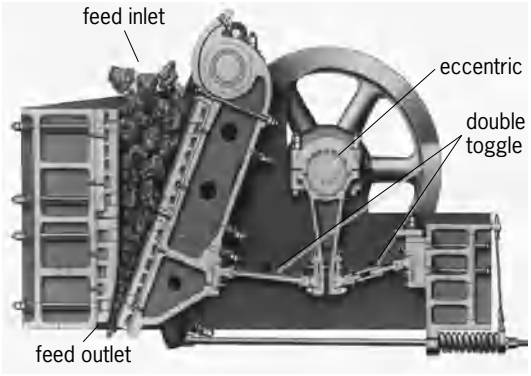


Fig. 1. Blake-type jaw crusher. (*Allis Chalmers Co.*)

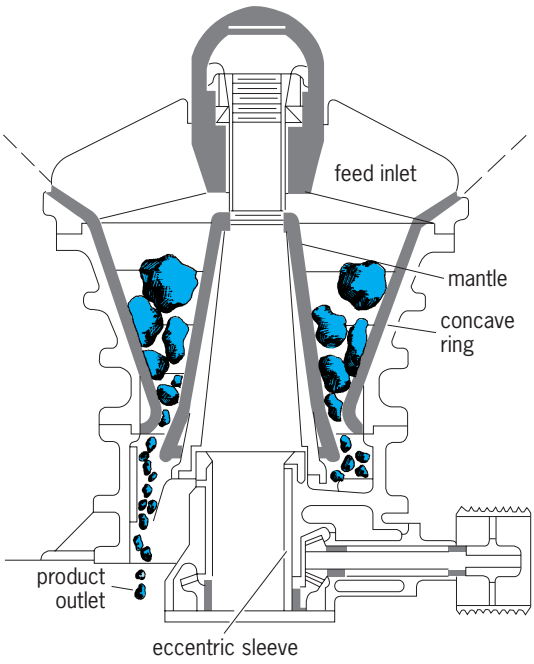


Fig. 2. Gates-type gyratory crusher. (*Allis Chalmers Co.*)

or eccentric and is generally built in smaller sizes.

The Gates gyratory crusher has a cone or mantle that does not rotate but is moved eccentrically by the lower bearing sleeve (**Fig. 2**). A 42- by 134-in. (107-by 340-cm) crusher has a capacity of 850 tons/h (770 metric tons/h), crushing rock from 27 to 8 in. (69 to 20 cm).

The Symons cone crusher also has a gyratory motion, but has a much flatter mantle or cone than does the gyratory crusher (**Fig. 3**). The top bowl is spring-mounted. It is used as a primary or secondary crusher.

Angle of nip is the largest angle that will just grip a lump between the gyratory mantle and ring, between the smooth jaws of a jaw crusher, or between the pair of smooth rolls. Depending on the material, it is 8–48°, but 32° is commonly used.

Secondary crushers. The single-roll crusher and the double-roll crusher have teeth on the roll surface and are used mainly for coal (**Fig. 4**). Single-roll crushers 36 in. (91 cm) in diameter by 54 in. (137 cm) long have a capacity of 275 tons/h (250 metric tons/h) crushing run-of-mine coal to 1 1/4-in. (32-mm) ring size. Smooth rolls without teeth are sometimes used

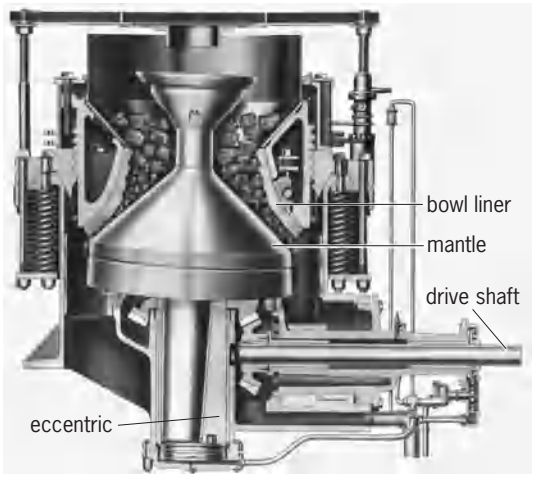


Fig. 3. Symons cone crusher. (*Nordberg Co.*)

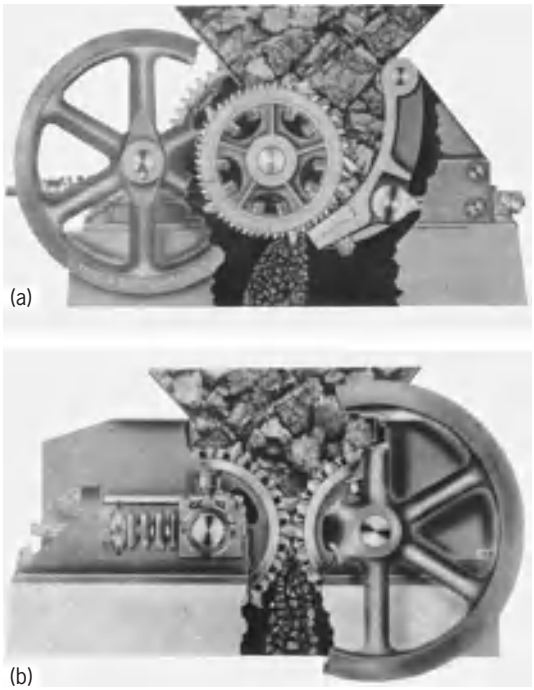


Fig. 4. Types of roll crusher. (*a*) Single-roll. (*b*) Double-roll. (*Link Belt Co.*)

for crushing ores and rocks.

The hammer crusher is the type of secondary crusher most generally used for ore, rock, and coal (**Fig. 5**). The reversible hammer mill can run alternately in either direction, thus wearing both sides of the hammers (**Fig. 6***a*). A hammer mill 42 in. (107 cm) in diameter by 82 in. (208 cm) long crushes 300 tons/h (270 metric tons/h) of coal to $^1/_4$ in. (6 mm) for coke oven feed. The ring coal crusher is a modification of the hammer mill using rings instead of hammers (Fig. 6*b*).

Pulverizers. In open-circuit pulverizing, the material passes through the pulverizer once with no removal of fines or recirculation.

In closed-circuit pulverizing, the material discharged from the pulverizer is passed through an external classifier where the finished product is removed and the oversize is returned to the pulverizer for further grinding.

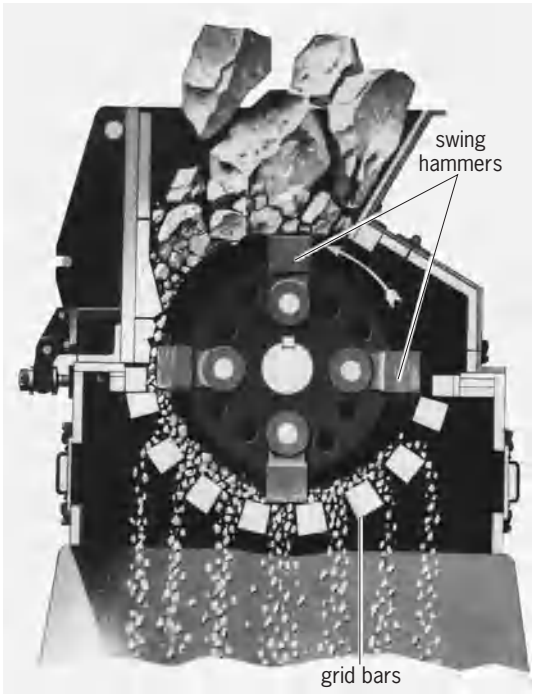


Fig. 5. Hammer crusher, a secondary crusher for ore, rock, and coal. (*Jeffrey Co.*)

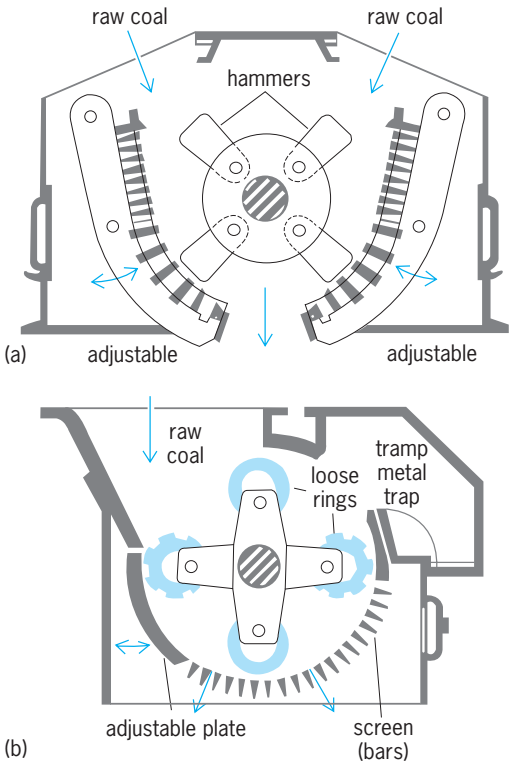


Fig. 6. Two examples of coal crusher. (*a*) Hammer mill type. (*b*) Ring type. (*Penna Crusher*)

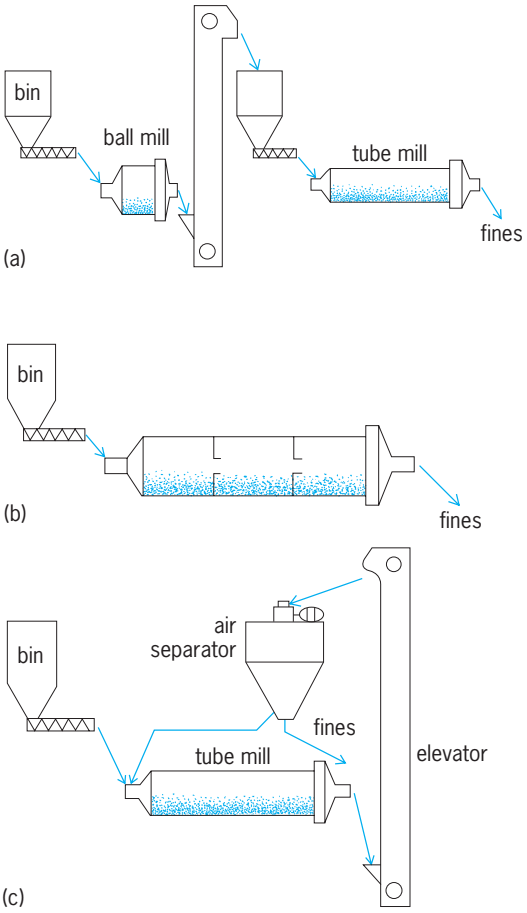


Fig. 7. Pulverizers. (*a*) Open-circuit type with preliminary ball mill stage followed by tube mill. (*b*) Open-circuit three-compartment tube mill. (*c*) Closed-circuit tube mill with an air separator.

Ball and tube mills, rod mills, hammer mills, and attrition mills are pulverizers operating by impact and attrition. In ball race and roller pulverizers, crushing and attrition are used.

Buhrstone pulverizer. The Buhrstone pulverizer is one of the oldest forms of pulverizers; a few still operate in flour and feed mills. The pulverizing is done between a rotating lower stone and a floating upper stone, which is restrained from rotating. Modern pulverizers use tumbling or rolling action instead of the grinding action of such earlier mills. For example, smooth double rolls are used for pulverizing flour and other materials. *See* BUHRSTONE MILL.

Tumbling pulverizers. There are two principal classes of tumbling pulverizer. Both consist of a horizontal rotating cylinder containing a charge of tumbling or cascading steel balls, pebbles, or rods. If the length of the cylinder is less than 2–3 diameters, it is a ball mill. If the length is greater than 3 diameters, it is a tube mill. Both types are extensively used in ore and rock pulverizing where the material is abrasive, as it is easy to add balls during operation. *See* TUMBLING MILL.

Ball mills charged with large steel balls 2–3 in. (5–8 cm) in diameter are often used as preliminary pulverizers in series with a tube mill in open circuit (**Fig. 7***a*). The Krupp mill is a preliminary ball mill which discharges the material through a screen which forms the cylindrical surface.

Pebble mills use balls and liners of quartz for application where contamination of the product with iron would be detrimental. For products coarser than 6 mesh, steel rods are sometimes used instead of balls. *See* PEBBLE MILL.

In some cement plants and many mining operations, ball, tube, rod, and pebble mills operate wet, water being added with the feed and the product being discharged as a slurry. Rake or cone type classifiers are used in closed circuits that employ wet grinding.

Tube mills for fine pulverization are used extensively in open circuit, or as compartment mills, or as closed-circuit mills (Fig. 7). A three-compartment, grate-discharge tube mill has larger balls in the first, medium-size balls in the second, and smaller balls in the third compartment (**Fig. 8**).

Cascade pulverizers are another form of tumbling pulverizer; they use the large lumps to do the pulverizing. The Aerfall pulverizer is built in diameters up to 16 ft (5 m). Feed is 12- to 18-in. (30- to 46-cm) lumps and the unit is airswept with an external classifier. A few steel balls are sometimes used but not over $2^1/_2$% of the volume.

The airswept conical ball mill is used in closed circuit with an external air classifier (**Fig. 9**). The

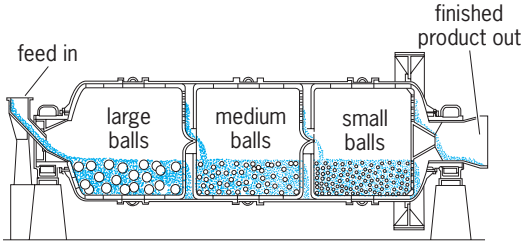


Fig. 8. Three-compartment tube mill pulverizer, containing three sizes of balls. (*Hardinge Co., Inc.*)

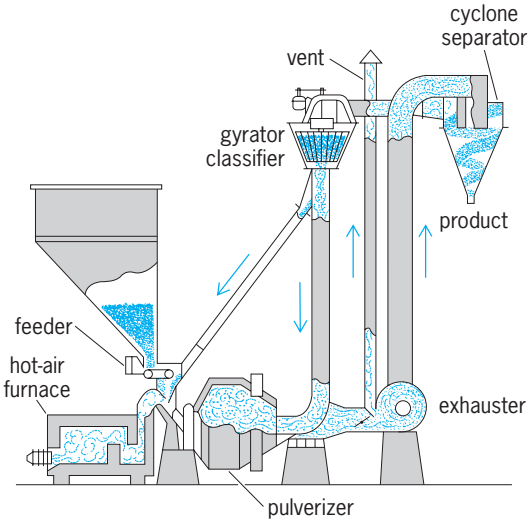


Fig. 9. Conical ball mill pulverizer in closed circuit with classifier. (*Hardinge Co., Inc.*)

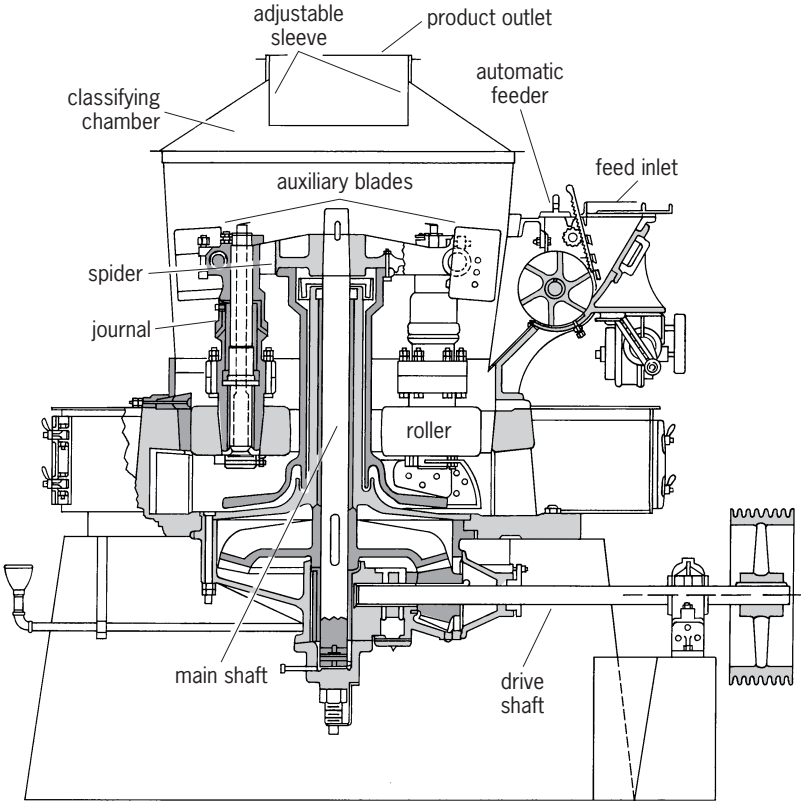


Fig. 10. A pulverizer of the airswept pendulum type. (*Raymond Pulverizer Division, Combustion Engineering Co.*)

conical shape of the mill classifies the balls, keeping the large balls at the feed end, where they are needed to crush the larger lumps, and the smaller balls at the small discharge end where the material is finer. This conical ball mill is also used extensively in open circuit in rocks and ores.

Roller pulverizers. The airswept pendulum roller pulverizer has rollers pivoted from a rotating spider on a vertical shaft; the centrifugal force of the rollers supplies the pressure between rollers and ring. Hot air enters below the rollers and picks up the fines. The rotating rollers and spider provide a centrifugal classifying action which returns the oversize to the pulverizing zone. For higher fineness a rotating classifier is added in the top housing (**Fig. 10**).

The ball race pulverizer uses large steel balls between races (**Fig. 11**). The bottom race is fixed, the intermediate race is rotated through driving lugs, and the top races are restrained from rotating by stops and are loaded by the springs. The material flows by centrifugal force from the center outward through the $12^1/_4$-in. (31-cm) and $9^1/_2$-in. (23.5-cm) balls. A ball race unit in closed circuit with an air separator is used for pulverizing cement rock.

Screens. Material is separated into desired size ranges by screening. For products 2 in. (5 cm) or larger, an open-end perforated cylindrical rotary screen is used. It is on a slight inclination, and the undersize goes through the holes and the oversize passes out through the open lower end.

For finer products from 2 in. (5 cm) down to fine

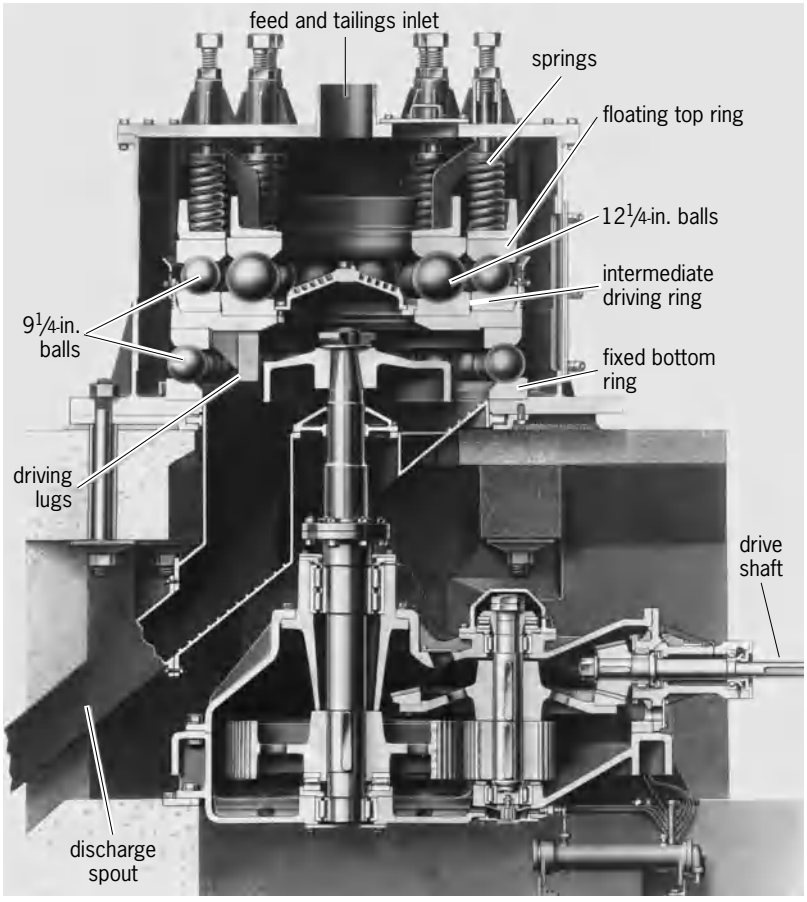


Fig. 11. Ball race rock pulverizer for use in closed circuit. (*Babcock and Wilcox Co.*)

powder passing through a 200-mesh sieve, shaking or vibrating screens of woven wire are used.

Shaking screens have either an eccentric drive or an unbalanced rotating weight to produce the shaking. Vibrating screens have electric vibrators (**Fig. 12**). They are often hung by rods and springs from overhead steel, giving complete freedom to vibrate. In some applications the screen is double decked, thus producing three sizes of product.

Direct-fired pulverizers for coal. Most modern large coal-fired furnaces are fed directly by the pulverizer, the coal rate being regulated by the rate of pulverizing. Air at 300–600°F (150–315°C) is supplied to the pulverizer to dry and preheat the coal. In a bowl-type medium-speed pulverizer, the springs press the pivoted stationary rolls against the rotating bowl grinding ring, crushing the coal between them (**Fig. 13**).

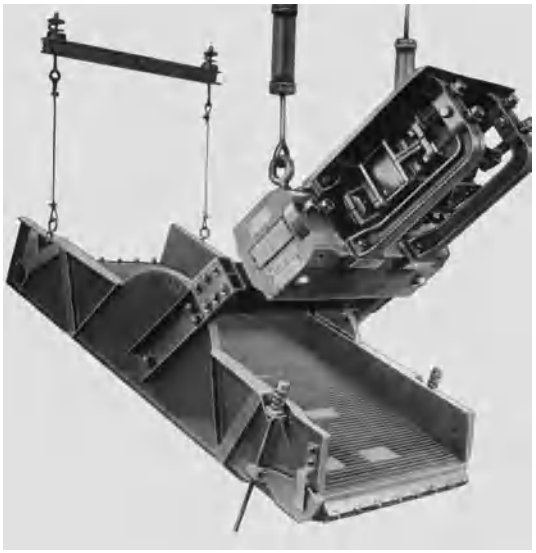


Fig. 12. Vibrating screen classifier. (*Jeffrey Co.*)

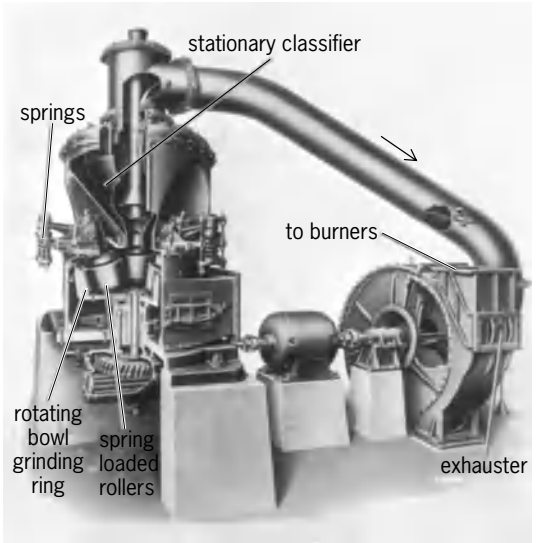


Fig. 13. Medium-speed bowl-mill coal pulverizer, which feeds burners. (*Combustion Engineering Co.*)

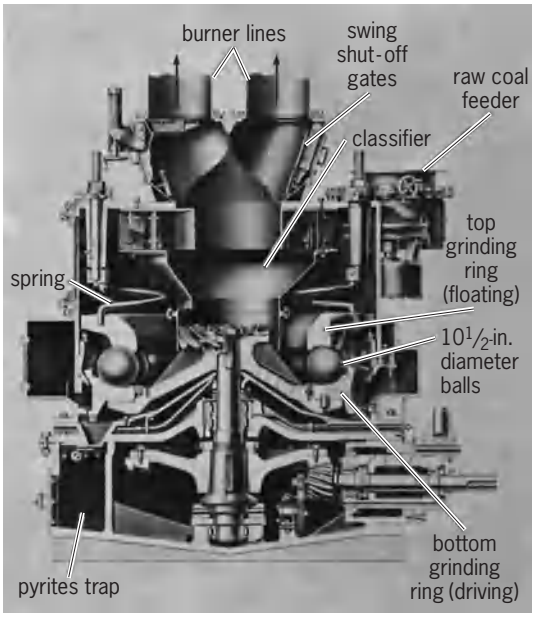


Fig. 14. Medium-speed ball race coal pulverizer, operated under pressure. 10½ in. = 27 cm. (*Babcock and Wilcox Co.*)

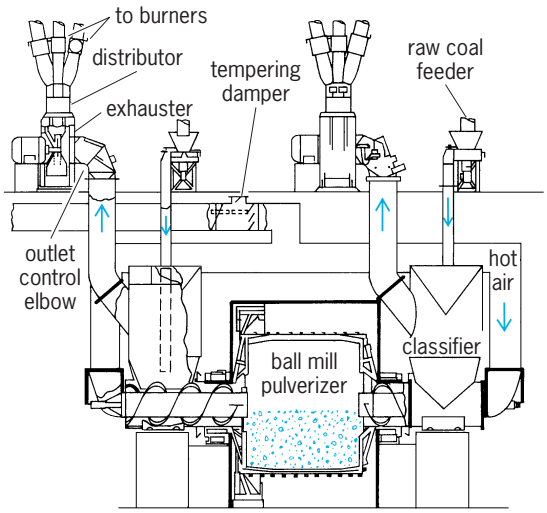


Fig. 15. Slow-speed coal pulverizer of the ball mill type; classifier sorts out oversize. (*Foster Wheeler Corp.*)

The self-contained stationary cone classifier returns the oversize to the pulverizing zone.

The medium-speed pulverizer of the ball race type uses closely spaced 18-in.-diameter (46-cm) balls between a lower rotating race and a floating top race (**Fig. 14**). The single coil springs restrain the top race and also apply the pressure needed. A stationary cone classifier is used. This pulverizer operates under pressure, using a blower on the clean inlet air instead of an exhauster on the coal-laden outlet air.

Slow-speed pulverizers of the ball mill type have classifiers and return the oversize for regrinding on both ends (**Fig. 15**).

High-speed pulverizers usually incorporate multiple stages such as a preliminary hammer mill and a secondary pulverizing stage. A built-in exhaust fan

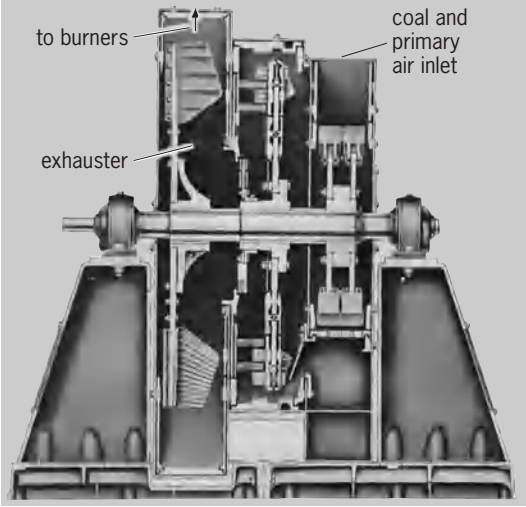


Fig. 16. High-speed coal pulverizer. (*Riley Stoker Corp.*)

exhausts the fuel and air from the unit (**Fig. 16**). *See* BALL-AND-RACE-TYPE PULVERIZER. Ralph M. Hardgrove

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Crustacea

An extremely diverse, species-rich group of arthropods that have inhabited marine environments since the beginning of the Cambrian Period. In some ways, they may be thought of as the marine equivalents of the insects. Within the marine realm the crustaceans occupy as varied a spectrum of habitats as the insects inhabit on land. The Crustacea currently are regarded as a subphylum in the monophyletic phylum Arthropoda, although debate about this rank continues among carcinologists.

The Crustacea are one of the most complex animal groups to define because of their great diversity of structure, habit, habitat, and development. No one character or generalization will apply equally well to all. The adults of some aberrant or highly specialized parasitic species are not recognizable as crustaceans, and evidence of their kinship rests on knowledge of their life histories. Virtually every form of marine animal life, as well as a number of freshwater and even terrestrial animals, has its crustacean parasites.

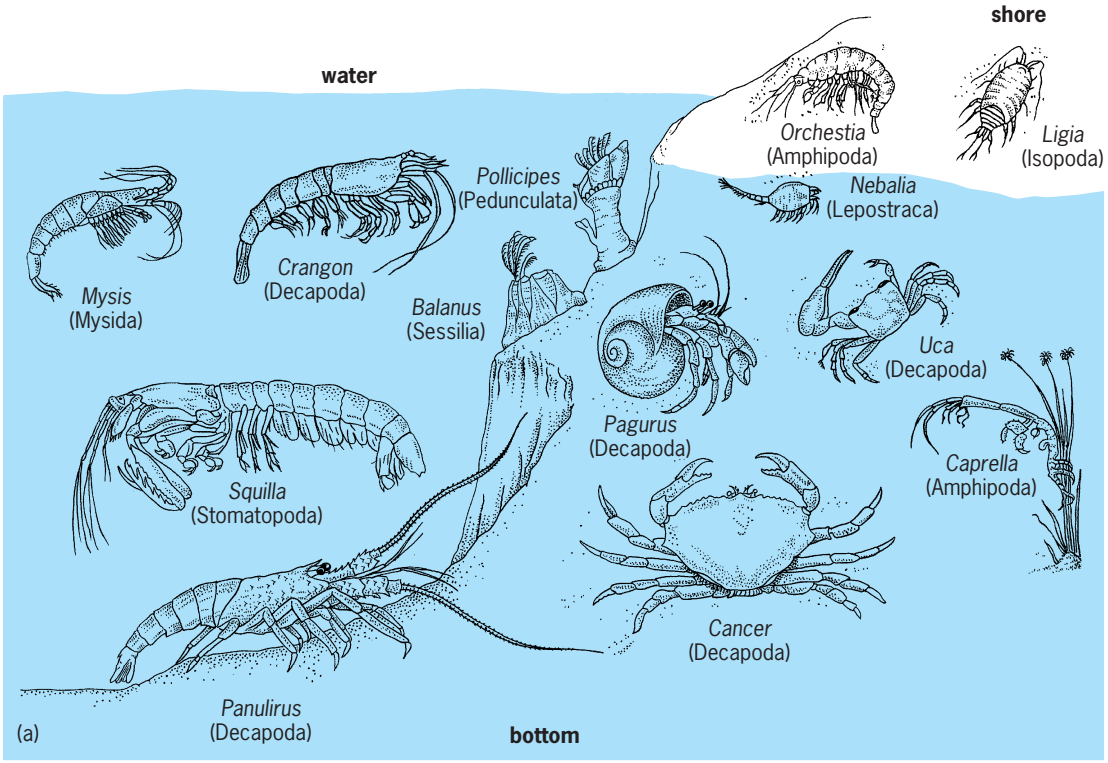
Characterization

Unlike the other arthropod subphyla, the Crustacea exhibit not one but several body plans (*Baupläne*) indicative of higher taxa within the subphylum. In addition to the chitinous exoskeleton—which may vary from thin, almost membranous and transparent to thick and impregnated with calcium carbonate—all crustaceans, unless unduly modified for a parasitic lifestyle, have a segmented body differentiated into dorsal tergites and ventral sternites, with each somite bearing a pair of appendages. Cephalic segmentation varies little among the crustacean classes, consisting of five, often fused somites with paired antennules, antennae, mandibles, maxillules, and maxillae on somites 1–5, respectively. The eyes are borne on an acronal region considered by some to be presegmental and by others to represent a sixth cephalic somite. Traditionally, the postcephalic body "trunk" was divided into a thorax (with a variable number of somites in lower, or less morphologically advanced, crustaceans and eight in higher taxa) and an abdomen (also with a variable number of somites in lower crustaceans but only six or seven in higher taxa). However, recent advances in developmental genetics have broadened the interpretation of trunk segmentation. Based at least in part on *Hox* gene expression patterns, the crustacean trunk now is delineated by the presence or the anteriormost expression of the *Hox* gene *Abd-B*. The trunk is divided into a thorax, defined as the region just posterior to the cephalon, which is differentiated from posterior regions by a structurally distinct set of appendages, and which is anterior to or coincident with the anteriormost expression of *Abd-B*; and a pleon, defined as the region posterior to the thorax, which is differentiated by a structurally distinct set of limbs (which may be reduced or absent), and which is typically posterior to the gonopores (and the postulated anteriormost expression of *Abd-B*.) Posterior to the trunk is the abdomen, defined as the region posterior to the trunk that lacks appendages; it is posterior to the expression of *Abd-B* and exhibits no expression of *Hox* genes. *See* DEVELOPMENTAL GENETICS.

Distribution and Ecology

The Crustacea are ubiquitous. They live at almost all depths and levels of the sea, in freshwaters at elevations up to 12,000 ft (3658 m), in melted snow water, in the deepest sea abysses more than 6 mi (9 km) down, and in waters of 0°C (32°F) temperature. They also live on land, in trees, and in mountains, although most of these species must descend to salt water areas again to spawn their young. Some live in strongly alkaline waters and others in salt water that is at the saturation point; still others exist in hot springs and hydrothermal vents with temperatures in excess of 55°C (131°F). Species such as shrimps, prawns, crabs, or lobsters (**Fig. 1**) are familiar. However, there are many more with vernacular names such as water fleas, beach fleas, sand hoppers, fish lice, wood lice, sow bugs, pill bugs, barnacles, scuds, slaters, and krill or whale food, and many with no vernacular names at all.

All crustaceans of sufficient size, if not directly eaten by people somewhere in the world, are used as bait for food. Only a very few, if properly prepared, have proved to be poisonous. Myriads of small fry are the sustenance of larger aquatic animals, chiefly fish and whales. Crustaceans come in all sizes, from minute and microscopic (less than 1 millimeter in total length) to very large with spans in excess of 12 ft (3.7 m) from tip to tip of the laterally extended legs. Individuals of the giant xanthid crab of Australia have been recorded weighing 20 to 30 lb (9 to



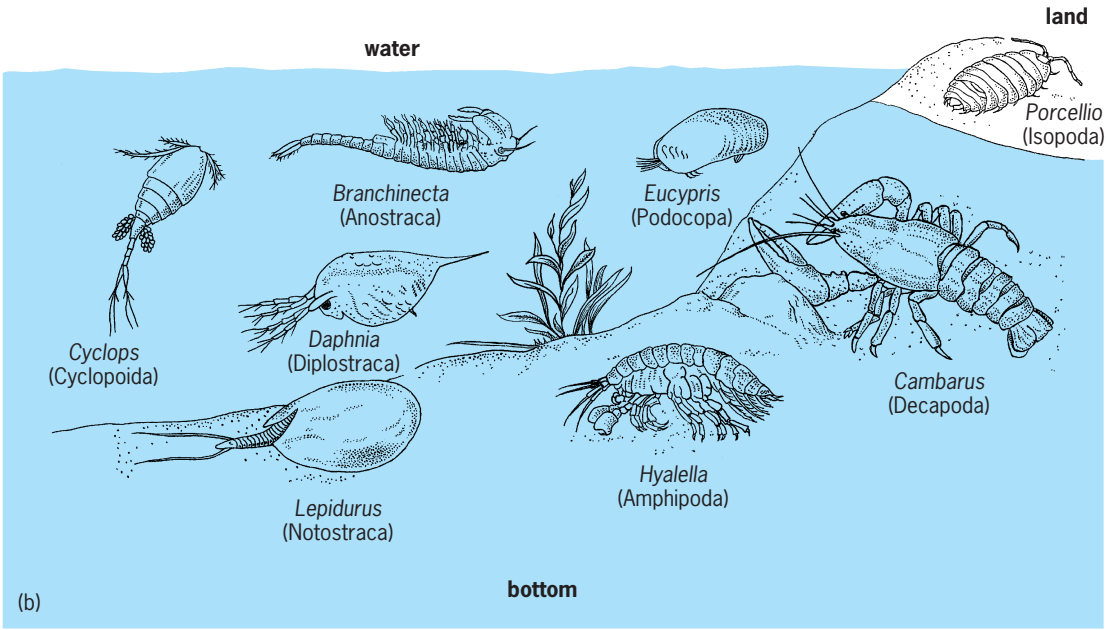


Fig. 1. Some crustaceans in their respective habitats, mostly reduced but not to same scale. Orders are in parentheses. (*a*) Marine forms and common names: *Pollicipes*, goose barnacle; *Balanus*, acorn barnacle; *Mysis*, opossum shrimp; *Orchestia*, sand hopper; *Caprella*, skeleton shrimp; *Crangon*, shrimp; *Panulirus*, spiny rock lobster; *Pagurus*, hermit crab; *Cancer*, edible crab; *Uca*, fiddler crab; *Squilla*, mantis shrimp. (*b*) Freshwater forms and common names if recognized. *Branchinecta*, fairy shrimp; *Daphnia*, water flea; *Eucypris*, seed or mussel shrimp; *Lepidurus*, tadpole shrimp; *Cyclops*, copepod; *Porcellio*, sow bug (on land); *Cambarus*, crayfish. (*After T. I. Storer et al., General Zoology, 6th ed., McGraw-Hill, 1979*)

14 kg), and the American lobster, the heaviest crustacean known, can weigh 44 lb (20 kg).

Most crustaceans are omnivorous and essentially scavengers. Many are filter feeders and screen particulate life, plankton, and organic detritus from the waters in which they live; others are largely carnivorous; and some are vegetarian. Among the vegetarians are the grazers of the ocean meadows, which convert the microscopic plant life (diatoms) into flesh and food for larger animals, which in turn are harvested as food for humans. *See* FEEDING MECHANISMS (INVERTEBRATE).

Taxonomy

Although disagreement persists regarding elements of the hierarchical classification of the Crustacea, in recent years a consensus, for the most part, has been reached. Still contentious is the class Maxillopoda, for which no clearly defined, single "good" character is shared by the various assigned taxa. Nonetheless, until such time as molecular, fossil, and morphological studies on members of the class provide some level of compromise, it seems in the best interest of nomenclatorial stability to continue to recognize the class Maxillopoda. *See* MAXILLOPODA.

The current crustacean classification scheme (based on Martin and Davis, 2001) is as follows (see separate articles on many of these texa):

Subphylum Crustacea
- Class Branchiopoda
  - Subclass Sarsostraca
    - Order Anostraca
  - Subclass Phyllopoda
    - Order: Notostraca
      - Diplostraca
- Class Remipedia
  - Order Nectiopoda
- Class Cephalocarida
  - Order Brachypoda
- Class Maxillopoda
  - Subclass Thecostraca
    - Infraclass Fascetotecta
    - Infraclass Ascothoracida
      - Order: Laurida
        - Dendrogastrida
    - Infraclass Cirripeda
      - Superorder Acrothoracica
        - Order: Pygophora
          - Apygophora
      - Superorder Rhizocephala
        - Order: Kentrogonida
          - Akentrogonida
      - Superorder Thoracica
        - Order: Pedunculate
          - Sessilia
  - Subclass Tantulocarida
  - Subclass Branchiura
    - Order Arguloida
  - Subclass Pentastomida
    - Order: Cephalobaenida
      - Porocephalida
  - Subclass Mystacocarida
    - Order Mystacocaridida
  - Subclass Copepoda
    - Infraclass Progymnoplea
      - Order Platycopioida
    - Infraclass Neocopepoda
      - Superorder Gymnoplea
        - Order Calanoida
      - Superorder Podoplea
        - Order: Misophrioida
          - Cyclopoida
          - Gelyelloida
          - Mormonilloida
          - Harpacticoida
          - Poecilostomatoida
          - Siphonostomatoida
          - Monstrilloida
- Class Ostracoda
  - Subclass Myodocopa
    - Order: Myodocopida
      - Halocyprida
  - Subclass Podocopa
    - Order: Platycopida
      - Podocopida
- Class Malacostraca
  - Subclass Phyllocarida
    - Order Leptostraca
  - Subclass Hoplocarida
    - Order Stomatopoda
  - Subclass Eumalacostraca
- Superorder Syncarida
  - Order: Bathynellacea
    - Anaspidacea
- Superorder Peracarida
  - Order: Spelaeogriphacea
    - Thermobaenacea
    - Lophogastrida
    - Mysida
    - Mictacea
    - Amphipoda
    - Isopoda
    - Tanaidacea
    - Cumacea
- Superorder Eucarida
  - Order: Euphausiacea
    - Amphionidacea
    - Decapoda

One group of parasites now acknowledged as probably crustacean is the Pentastomida. Until quite recently, the Pentastomida had been classified as a distinct phylum, all adult members of which are parasites of the respiratory tracts of vertebrates. However, molecular and spermatological evidence of a phylogenetic relationship between pentastomids and branchiurans accumulated over the past quarter-century—with additional supporting evidence from similarities in embryogenesis and details of cuticular and nervous system structures—has resulted in the acceptance of this very distinctive group of parasites as crustacean in origin. The primary larvae are autoinfective to the definitive host, but they can also migrate to the host's gut and pass out with the feces. Many pentastomids require an intermediate host, which may be any kind of invertebrate. Larvae bore through the gut wall of the intermediate host, where they undergo further development to the infective stage. Once the intermediate host is consumed by a definitive host, the parasite makes its way from the host's stomach to its final destination in the lungs, nasal passages, or bronchi.

If this current phylogenetic assessment is accurate, crustacean morphological diversity and extremes in lifestyle are taken to an even greater level, with representatives adapted to life in the respiratory passages of crocodilians, reindeer, and lions. However,

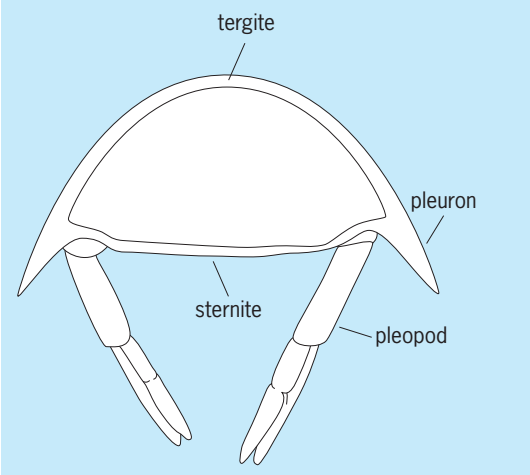


Fig. 2. Diagram of the fifth pleonal somite of the crayfish, with its paired fifth pleopods.

the placement of the Pentastomoida within the Crustacea is not universally accepted. The discovery of fossils from the Middle Cambrian limestone that are very similar to extant pentastomids would appear to dispute this classification. *See* PARASITOLOGY; PENTASTOMIDA.

General Morphology

The true body segments, the somites or metameres, are usually somewhat compressed or depressed (**Fig. 2**). Rarely are they circular rings. The somite's dorsum, or upper side, is the tergum or tergite; the lower or ventral surface, the sternum or sternite; and either lateral portion, the pleuron or epimeron. When prolonged or extended downward to overhang the attachment of the biramous appendages, of which one pair is a part of every typical crustacean somite, the structure is termed the epimeron. Between the embryonic anterior or prostomial element and the telson at the posterior end, the linear series of somites making up the body of a crustacean are more or less distinctly organized into three or four regions, or tagmata: the head (cephalon), thorax, and abdomen in lower crustaceans (for example, Branchiopoda, Cephalocarida, Maxillopoda) and head, thorax, pleon, and abdomen in malacostracans. The somites are variously fused with one another in diagnostic combinations in different groups of the Crustacea.

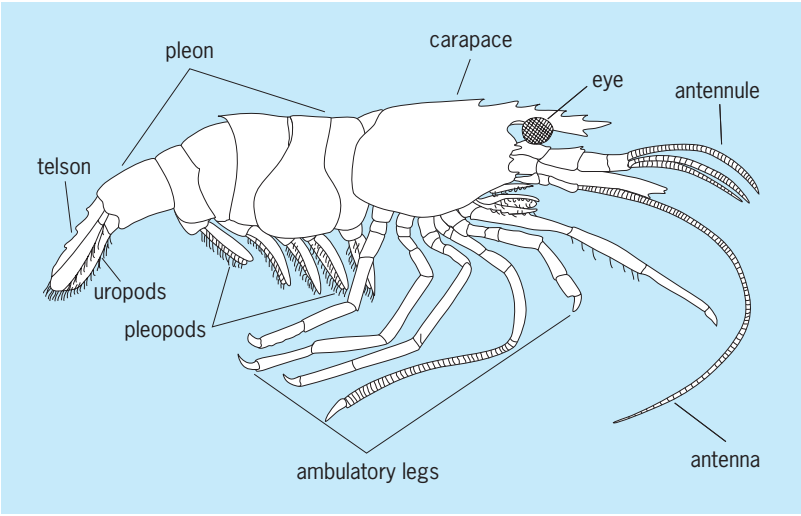


Fig. 3. External morphology of a shrimp. (*After P. A. McLaughlin, Comparative Morphology of Recent Crustacea, W. H. Freeman, 1980*)

Body. The head region of the simplest type is found in the characteristic crustacean larva, the nauplius. When the nauplius hatches, it has only three pairs of appendages—the antennules, antennae, and mandibles—which correspond to the three fused somites of the primary head of the lower Crustacea. In the nauplius, there is a single median eye (known as the nauplius eye on the head that persists to become the sole visual organ in many of the lower Crustacea. In some it disappears; occasionally it remains as a vestige in species with paired compound eyes.

In more advanced forms, and at later developmental stages, two additional somites bearing the maxillules and maxillae become coalesced with the protocephalon to form the most frequently encountered crustacean head, or cephalon. Thoracic somites frequently are fused with the cephalon to form a cephalothorax. A dorsal shield, or carapace, of variable length arises from the dorsum of the third cephalic somite and covers the cephalon and cephalothorax to a varying extent (**Fig. 3**). In some crustaceans the carapace develops in the form of a bivalved shell; in sessile barnacles it is modified as a fleshy mantle reinforced by calcareous plates. The carapace reaches its greatest development in the malacostracan Decapoda where, including the head, it fuses dorsally with the eight thoracic somites and extends down laterally over the attachment of the appendages to enclose the branchial chamber on either side of the cephalothorax.

The chitinous cuticle covering the crustacean body is its external skeleton (exoskeleton). Ingrowths for muscle attachment develop from this cuticular exoskeleton as apodemes. Further development of the apodemes results in the formation of the endophragmal skeleton in the higher Crustacea. The chitin is flexible at the joints, in foliaceous appendages, and throughout the exoskeletons of many small and soft-bodied species, but it is often thickened and stiffened in others. It becomes calcified in many species as a result of the deposition of lime salts, and can be rock-hard in some crabs.

Appendages. The paired appendages of the cephalon and thorax (**Fig. 4**) are typically biramous and consist of two branches: the endopod and exopod. These arise from the distal end of a peduncle, the protopod, which is composed of two segments: the coxa, articulated with the body, and the basis; occasionally a precoxa may also be present. Either the endopod or exopod or both may be reduced or absent, or both may be coalesced, as in the foliaceous limbs of lower Crustacea. The protopod frequently has inner lobes (endites) or outer lobes (exites) or branchial structures (epipods). The

endopod is definitely segmented in the higher Crustacea. The segments, which are usually present in full complement from the peduncular basis outward, are the ischium, merus, carpus, propodus, and dactyl. The endopods are variously modified to serve a number of functions and needs, such as sensory perception, respiration, locomotion, prehension (holding, seizing, or grasping) and manipulation of food, cleansing, defense, offense, reproduction, and sex recognition and attraction. If retained in the adult, exopods may remain leaf- or paddlelike, or become flagellated structures, facilitating swimming or aiding respiration. Pleonal appendages develop only in the Malacostraca; they are commonly used in locomotion and/or reproduction. The nonmalacostracan crustacean abdomen lacks appendages.

Digestive system. The digestive system is composed of a gut divided into three distinct regions and its accompanying glands and diverticula. Appendages adapted for feeding frequently include not only the mandibles, maxillules, and maxillae but also a number of thoracic appendages and occasionally the antennae.

In its simplest form, the gut consists of a foregut with an anteroventrally directed mouth, midgut, and hindgut with a posterior or posteroventrally directed anal orifice. The anterior part of the foregut often is esophageal in nature, whereas the posterior and larger portion usually consists of two parts. (In the anterior foregut of the Eucarida, a complex and elaborate grinding mechanism, the gastric mill, is developed). The posterior chamber is divided into dorsal and ventral filtering compartments for the straining of food. The midgut in most cases contains several ceca or diverticula, which produce digestive secretions or serve as organs for the absorption of food. When these ceca are present in considerable number, they become organized into a sizable gland, the midgut gland of higher Crustacea. The hindgut is typically short and terminates, with few exceptions, in a muscular anus on the underside of the telson. *See* DIGESTION (INVERTEBRATE).

Circulatory system. The crustacean circulatory system is, with few exceptions, an open system through which the blood or hemolymph flows freely throughout the hemocoelic cavity. Not all crustaceans have a heart, but in those that do, it is a simple single-chambered organ lying dorsally above the gut and enclosed in a pericardial sinus. The heart is perforated by openings, or ostia, that admit venous blood from the pericardial sinus. The heart may be elongated and tubular and extend through the greater part of the body, but generally it is a more compact organ. It pumps blood through an arterial system or through connecting sinuses or lacunae within the body tissues. In Crustacea lacking a heart, blood is moved by the muscular movements of the animal or its alimentary tract, circulating it through the body cavities or sinuses. The blood of malacostracan crustaceans is bluish because it contains the respiratory pigment hemocyanin. A few crustaceans have red blood due to the presence of hemoglobin. *See* HEART (INVERTEBRATE); RESPIRATORY PIGMENTS (INVERTEBRATE).

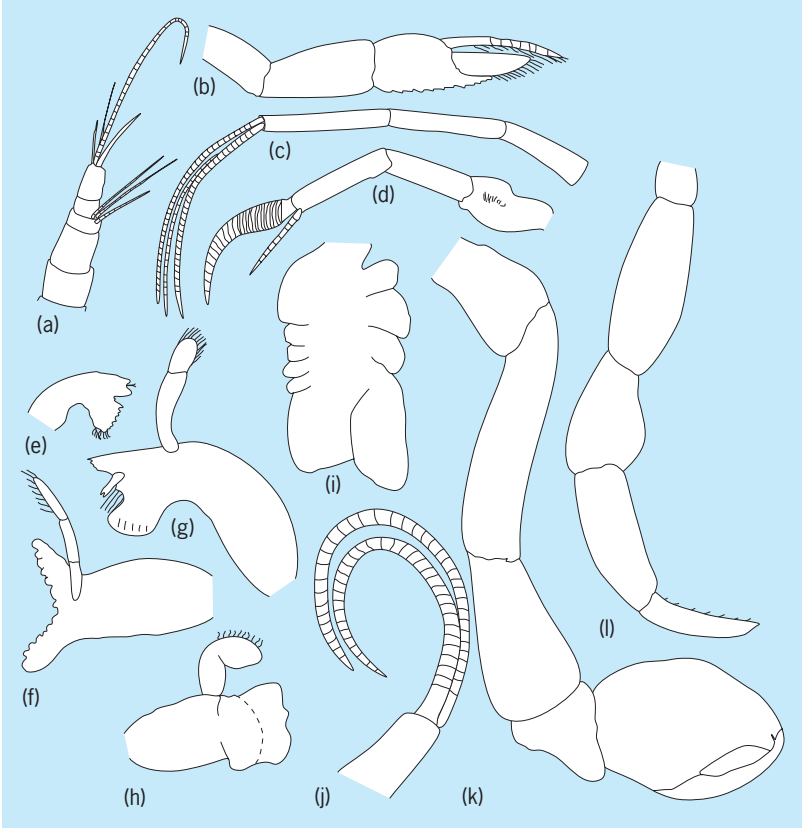


Fig. 4. Cephalic and thoracic appendages of representative crustaceans: (*a–d*) antennules; (*e–h*) mandibles; (*i–k*) thoracopods. (*a*) Brachypoda; (*b*) Leptostraca; (*c*, *d*) Decapoda; (*e*) Brachypoda; (*f*) Stomatopoda; (*g*) Mysida; (*h*) Decapoda; (*i*) Anostraca; (*j*) Sessilia; (*k*) Stomatopoda; (*l*) Decapoda. (*After P. A. McLaughlin, Comparative Morphology of Crustacean Appendages, in D. E. Bliss, ed., The Biology of Crustacea, vol. 2, 1982*)

Respiratory system. The majority of nonmalacostracan crustaceans have no specialized respiratory structures; respiration takes place across the integumental surfaces. However, in malacostracans oxygen uptake is accomplished principally by means of gills, although the general body surface, or special areas of it, also may participate. Some of the few species that have become more or less terrestrial in their habits have developed modifications of their branchial mechanism (such as villi, ridges, or water-retaining recesses), which when sufficiently moist enable them to breathe air. Some sow or pill bugs have special tracheal developments in their abdominal appendages for the same purpose.

Nervous system. As do other arthropods, crustaceans possess a dorsally located brain connected to a double ventral nerve cord by paired circumesophageal connectives. Both the brain and the ventral nerve cord are segmental structures, and each unit (neuromere) is joined by paired longitudinal connectives. This segmental aspect of the central nervous system is most evident in embryonic stages, and is generally obscured by condensation of the neuromeres into ganglia during later embryonic or larval development. In decapod embryos, the ventral nervous system consists of eight thoracic and eight pleonal neuromeres, whereas in many lower

crustaceans a much higher number of trunk neuromeres is present. *See* NERVOUS SYSTEM (INVERTEBRATE).

Organs of special sense. The organs of special sense are the eyes, antennules, and antennae. Crustacean eyes (photoreceptors) are of two types: sessile and stalked. Sessile eyes, found in numerous lower crustaceans and in larval stages of numerous stalk-eyed species, often are simple structures consisting of light-sensory elements, although some are weakly to moderately well-developed compound eyes. Stalked eyes are unique among the Crustacea; the stalks provide support and movement for the eyes, and most also contain optic ganglia and important endocrine organs. The crustacean compound eye contains from few to many structural units, or ommatidia.

The antennules and antennae are provided with a variety of sensory structures for the reception of chemical and mechanical stimuli. Specific structures on the antennules, called aesthetascs, act as organs of smell. Taste chemoreceptors are usually found on the mouthparts and dactyls of the pereopods (or pereiopods), or walking legs. Many of the hairs (setae) and bristles found on the crustacean body and appendages act as mechanoreceptors. Organs of balance (statocysts) are also present on basal segments of the antennules in many crustaceans, but on the uropods in most mysids. Although crustaceans may not hear in the accepted sense, they are sensitive to sound waves and vibrations; many have stridulating or sound-producing mechanisms on appendages or on somites overlapping one another on the carapace, which produce snapping, clicking, rasping, or rubbing noises. Crustacea have setae variously located on their bodies which are sensitive to various external stimuli. *See* CHEMORECEPTION; EYE (INVERTEBRATE); PHOTORECEPTION.

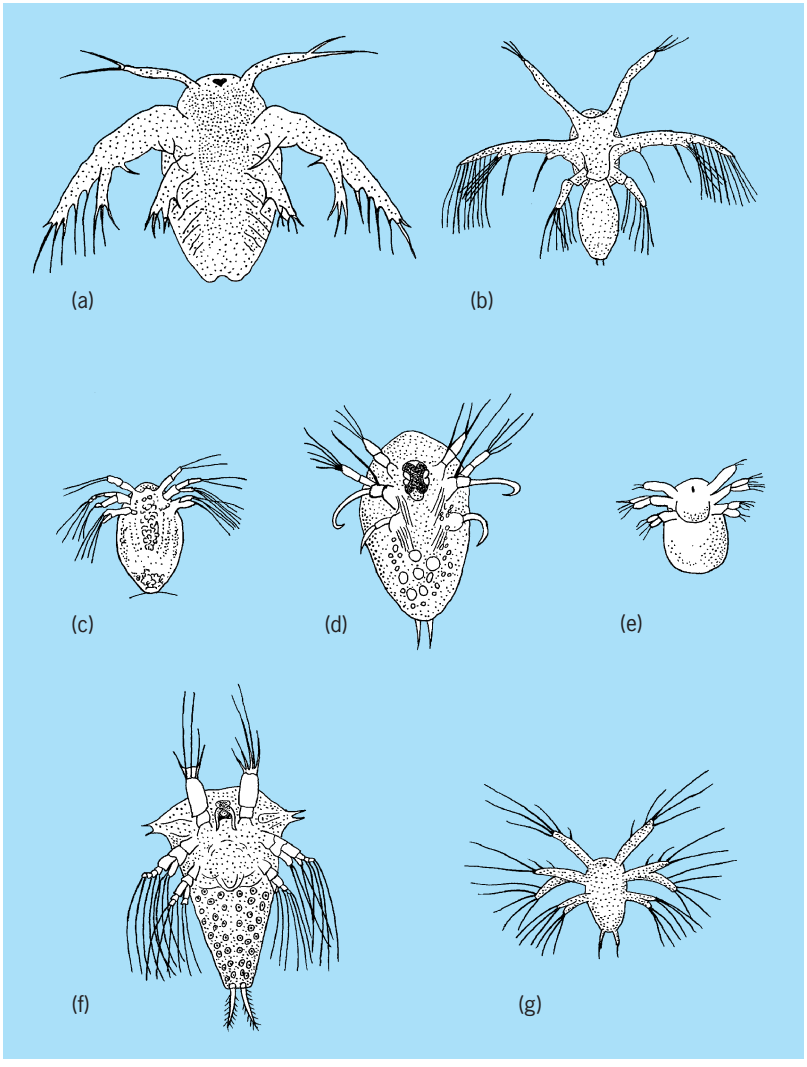


Fig. 5. Nauplii of (*a*) *Triops cancriformis* (Notostraca); (*b*) *Branchinecta occidentalis* (Anostraca); (*c*) *Lernaeocera branchialis* (Siphonostomatoida); (*d*) *Haemocera danae* (Monstrilloida); (*e*) *Meganyctiphanes norvegica* (Euphausiacea); (*f*) *Sacculina carcini* (Kentrogonida); (*g*) *Litopenaeus setiferus* (Decapoda). Enlarged but not to same scale. (*After R. E. Snodgrass, Crustacean Metamorphoses, vol. 131, no. 10, Smithson. Inst. Publ. 4260, 1956*)

Glands and glandular systems. Those glands recognized as definite excretory organs are the maxillary and antennal glands in the adult crustaceans. They are located in the cephalon, and rarely are both present at the same time. The antennal gland is best developed in the Malacostraca. It is often called the green gland and opens to the exterior on the underside of the coxal segment of the antennal peduncle. Many crustaceans, especially deep-sea forms, have phosphorescent or luminous organs. Barnacles and some other crustaceans have cement glands. This may also be true of tube-building species that construct their homes of extraneous materials. Species that produce encysted or drought-resistant eggs have other glands of special secretion. The X organ-sinus gland complex of the eyestalks is part of a neurosecretory system that produces hormones that control color change and pattern, molting cycle, oogenesis, and egg development within the ovary. *See* ENDOCRINE SYSTEM (INVERTEBRATE); NEUROSECRETION.

Reproductive system. The sexes are separate in most Crustacea and usually can be differentiated from each other by secondary sexual characters. Chief among these characters are the size and shape of the body, appendages, or both, and placement of the genital apertures. Hermaphroditism is the rule in the Cephalocarida, Remipedia, some ostracodes, and sessile Cirripedia (barnacles), in isolated cases in other crustaceans, and in certain parasitic forms. Parthenogenesis (eggs developing and hatching without prior fertilization) occurs frequently in some of the lower crustaceans such as the Branchiopoda and Ostracoda, which have what might be called an alternation of generations. The parthenogenetic generations (and there are usually very many of them in succession) alternate with a generation produced by fertilized eggs.

Protandric hermaphroditism, in which individuals reach maturity as males but subsequently become functional females, has been observed in some decapods. Protogyny, a condition in which the development is reversed, has been reported in tanaids and some carideans. Sexual differentiation is under control of the androgenic gland. This is a reversal of sex

in the individual or in the species that reach maturity as males but become female in their third year and function as females.

The eggs of most crustaceans are carried attached to the female until hatched. Some females develop brood pouches in which the young are retained for a time. A nutrient secretion, which sustains the young until they are released, is produced in some species having a brood chamber. Penaeid shrimp and a few of the lower Crustacea deposit their eggs in the medium in which they live, in some cases attaching them to aquatic vegetation.

Development. The nauplius larva has long been considered characteristic of Crustacea (**Fig. 5**). Its oval, unsegmented body has three pairs of appendages indicative of three embryonic somites. These are the antennules (which are uniramous in the nauplius), the biramous antennae, and mandibles. In addition to serving other functions such as sensation and feeding, all three pairs of appendages are the organs of locomotion of this free-swimming larva. There is a single dorsal, median, nauplius eye composed of three closely united, similar parts that form a relatively simple compound eye and a frontal organ. The integument of the upper surface of the nauplius body has the appearance of a dorsal shield, as in most crustaceans.

Typical as this first larval stage may be, it does not appear in all crustaceans. It is common in the lower forms, but among malacostracan crustaceans, free-living nauplius larvae occur in only the Euphausiacea and the decapod suborder Dendrobranchiata (penaeid shrimp). This larval form is replaced in many malacostracans by the so-called egg-nauplius, an embryonic stage characterized by an early germ band marking the ventral side of the embryo, which is attached to a relatively large amount of yolk and covered by a few extraembryonic cells. The anterior region is formed by the paired head lobes, which give rise to the lateral compound eyes and the protocerebrum. Posteriorly, these lobes are followed by the characteristic, well-developed antennules, antennae, and mandibles. The occurrence of an egg-nauplius is always correlated with a relatively yolky egg but is not necessarily correlated with direct development. Stomatopods and most decapods hatch as advanced planktonic larvae (zoeae) with at least one well-developed thoracic appendage. These larvae resemble the advanced stages of euphausiid and dendrobranchiate decapod larvae that hatch at the free-living naupliar stage.

Life histories vary from the simple to the complex within the different groups of Crustacea. Following the naupliar stage, whether it is passed in the egg or not, very remarkable larval stages follow in some groups the free-swimming cypris larvae of barnacles; the phyllosomae of the spiny lobsters; the pseudozoeal larvae of the mantis shrimps (Stomatopoda); and the zoeae and megalopae of crabs (**Fig. 6**).

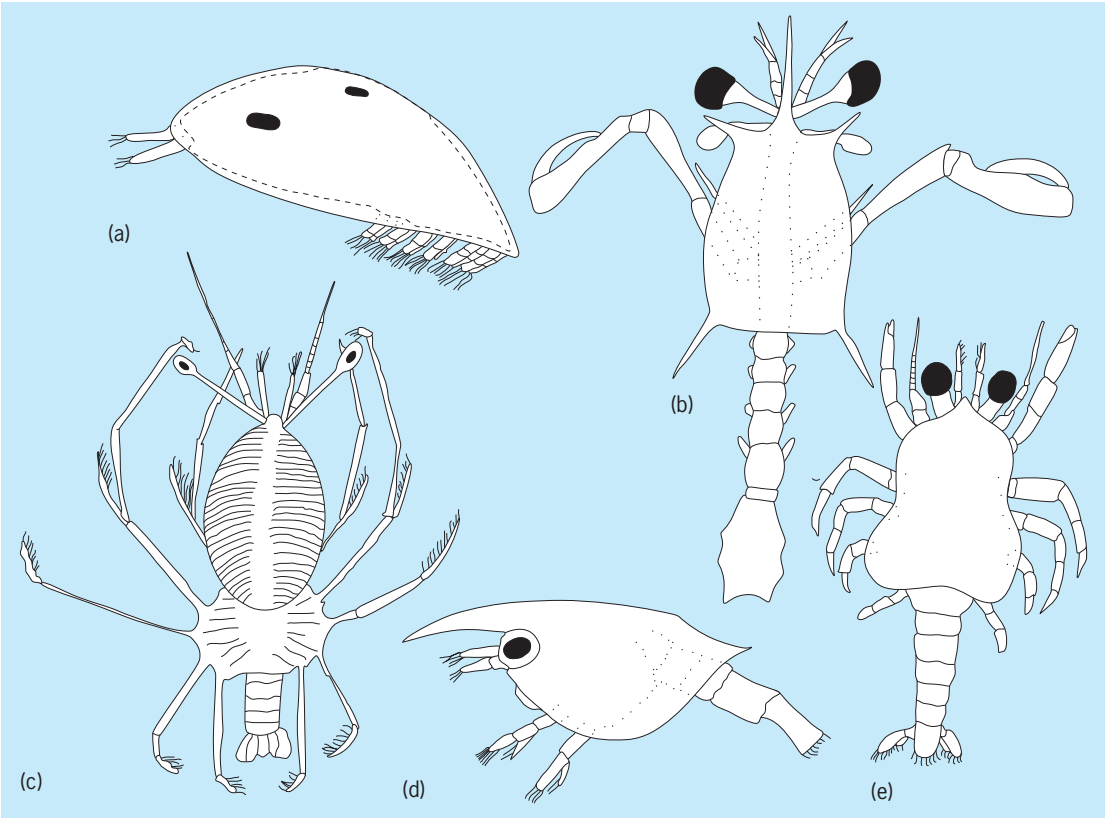


Fig. 6. Larval stages of representative crustaceans: (*a*) cyprid larva of Sessilia; (*b*) pseudolarva of Stomatopoda; (*c*) phyllosoma larva of Decapoda; (*d*) zoeal larva of Decapoda; (*e*) megalopa of Decapoda. (*After P. A. McLaughlin, Comparative Morphology of Recent Crustacea, W. H. Freeman, 1980*).

Physiological Mechanisms

Because of its tough, chitinous, or calcified exoskeleton, a crustacean does not usually grow without molting. Crustaceans have also developed the remarkable feature of autotomy and regeneration to minimize injury and hemorrhage from loss of limb through accident or to an enemy.

Molting (ecdysis). This process involves several steps: (1) preparation, which includes some degree of resorption of the old cuticle; (2) the formation of a new, temporarily soft and thin one within it; (3) the accumulation and storing of calcium in the midgut gland or as lenticular deposits (gastroliths), one of which occurs on either side of the foregut in certain forms such as crayfishes and lobsters. The preparatory period is less complicated in the thinly chitinous forms. The actual molt follows. The old shell or cuticle splits at strategic and predetermined places, permitting the crustacean within, already enclosed in new but still soft exoskeleton, to withdraw. Most likely, a temporary absorption of water enables the animal to split or crack its housing. Upon withdrawal of the entire animal, absorption of water again rapidly takes place with a pronounced increase in body size. The chitinous lining of the foregut and hindgut as well as the endophragmal skeleton are shed along with the exoskeleton. In the immediate postmolt period the crustacean is quite helpless and, if unsheltered, is at the mercy of its enemies. It is in the soft-shelled stage immediately after molting that the East Coast blue crab (*Callinectes*) becomes a much-sought seafood delicacy. The tender new cuticle is reinforced rapidly by the resorbed chitin, and hardened by whatever reserves of calcium the animal may have stored, supplemented and extended by the far more plentiful supplies in solution in the sea, which may be absorbed or ingested by the growing crustacean. Molting takes place quite frequently in the larval stages when growth is rapid but becomes less frequent as the animal ages. In many species there is a terminal molt at maturity. Hormones from the eyestalk play an important role in both initiating and inhibiting molting. *See* GASTROLITH.

Autotomy and regeneration. The mechanisms of autotomy and regeneration are developed in some crustaceans to minimize injury or loss to an enemy. When an appendage is broken, it is cast off or broken at the fracture or breaking plane (preferred breakage point or PBP). In lower crustaceans and even malacostracans such as amphipods and isopods, no preferred breakage point has been found, but its occurrence in brachyuran decapods is well documented. In most species the breaking point is localized at the joint between the basis and ischium, which is usually a nonfunctional articulation. Microcanals spanning the whole depth of the cuticle are more abundant here than in other regions of the appendage. No muscle develops through the breaking point; instead a specialized autotomizer muscle is inserted just proximal to it. The only nerve developing through the breaking point is locally tapered and, thus, weakly resistant. In addition, the blood vessels crossing the breaking point have a valve, preventing bleeding. Autotomy is well documented in some brachyuran decapods. Contraction of the autotomizer muscle pulls the anterior-basal part of the basis into the coxa. The mechanical resistance of the distal margin of the coxa against the anterior surface of the basi-ischium results in breaking of the integument between the basis and ischium; the nerve and blood vessels are then cut.

Crustacea also have the ability to regenerate lost parts. Although the regenerated parts are not always the same size as the original in the first molt after injury, increase in size in successive molts soon restores a lost limb to virtually its former appearance. During the regenerative stages among alpheid shrimps and lobsters, which generally exhibit marked dimorphism in the chelipeds when the major chela is lost, the minor one often may take on more or less the character of the former major chela. Then, the chela replacing the lost one assumes the habitus of the former minor chela during successive molts. *See* REGENERATION (BIOLOGY).

Geologic History

Exceptionally well-preserved arthropods have been discovered in Late Cambrian rocks. These arthropods are referred to as stem crustaceans, a group which is judged to have appeared at least by the Early Cambrian. The stem crustaceans are related to the true crustaceans, to which the former gave rise. Their morphology indicates that the Crustacea is most likely monophyletic, that is, derived from a single common ancestor. The oldest true crustaceans are ostracodes, dating from 570 million years ago (Ma), documented by species in several families. Thus, although the ostracodes are not the most primitive crustaceans, their remains constitute the oldest authentic record of the Crustacea. The more primitive classes (**Fig. 7**) have few hard parts likely to be preserved, and therefore they are far less well known. The Cephalocarida are known only from the Recent, and the Branchiopoda do not appear as fossils until the Early Devonian, about 395 Ma. The Remipedia first appear in the fossil record in the Carboniferous, about 330 Ma. The appearance of more primitive crustaceans later in geologic history, the reverse of what would be expected, is clearly more related to their preservation potential than to their evolutionary history. Furthermore, it is likely that the agnostids, a group of trilobitelike organisms, are more closely allied with the Crustacea than with the Trilobita. The more advanced classes, the Maxillopoda and the Malacostraca, appeared in the fossil record, respectively, in the Middle Cambrian, about 530 Ma, and the Late Devonian, about 370 Ma. *See* CEPHALOCARIDA; MALACOSTRACA.

The isopods and the decapods are among the most commonly known crustaceans. Isopods inhabit marine, freshwater, and terrestrial habitats and are known from the fossil record as far back as the Triassic, 230 Ma. The common pill bug represents one of

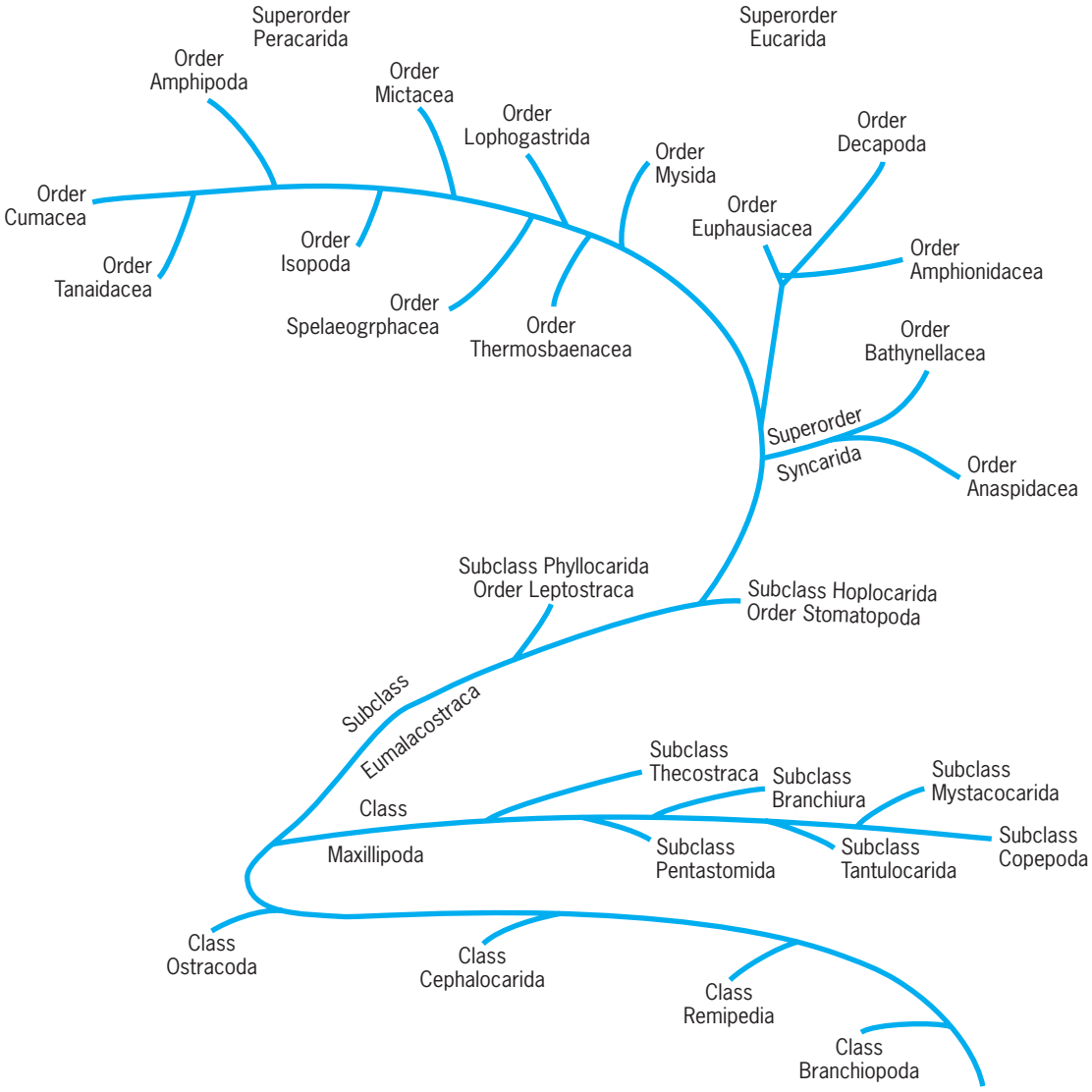


Fig. 7. Chart of the classification of Crustacea; no indication of phylogenetic relationships is implied. (*Modified from D. E. Bliss, ed., The Biology of Crustacea, vols. 1–10, Academic Press, 1982–1985*)

the few totally terrestrial isopods (Fig. 1). The decapod crustaceans, including the modern shrimp, crabs, and lobsters (Fig. 1), trace their beginnings to a lobsterlike form in the Late Devonian, 350 Ma (**Fig. 8**). Thus, the geologic history of the crustaceans is robust and diverse, and the marine environment, in particular, has been home to this group since the early part of the Paleozoic Era. *See* DECAPODA (CRUSTACEA); ISOPODA.

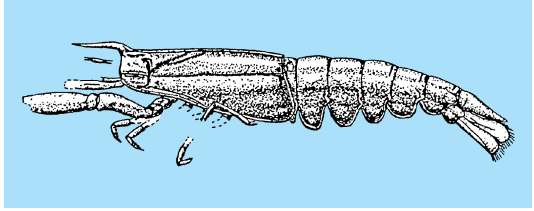


Fig. 8. Late Devonian malacostracan *Palaeopalaemon newberryi* Whitfield, considered to be the earliest decapod crustacean. (*After F. R. Schram, R. M. Feldmann, and M. J. Copeland, J. Paleontol., 52:1381, 1978*)

Waldo L. Schmitt; Patsy A. McLaughlin;
Rodney M. Feldmann

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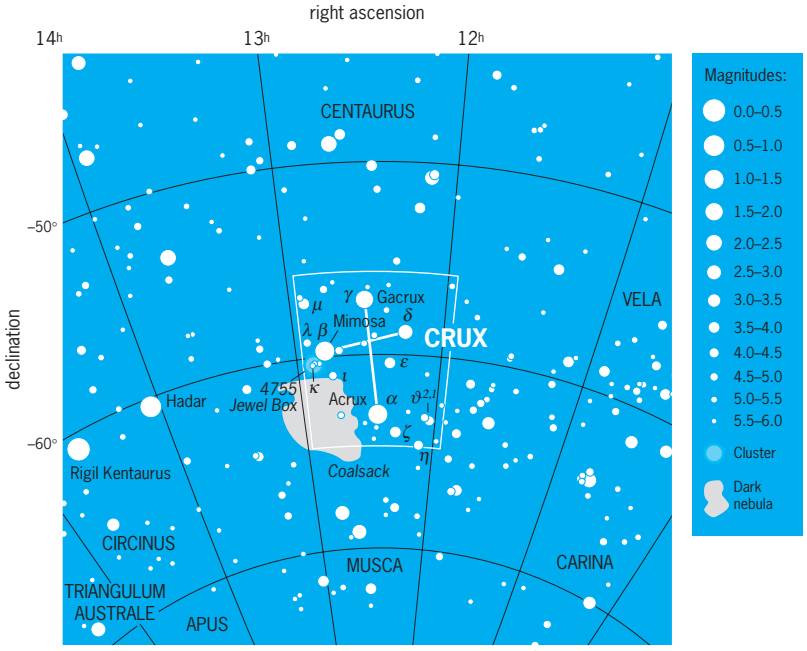
Crux

The Southern Cross, a small southern constellation. Because of the absence of a star at the center of the cross, this constellation really looks like a kite (see **illustration**). To its side, the Coalsack is a giant dark silhouette of dust against the Milky Way Galaxy. In it the colorful open star cluster known as the Jewel Box is found. It contains dozens of hot, bright, blue stars plus a prominent, central red supergiant. *See* STAR CLUSTERS.



Modern boundaries of the constellation Crux, the Southern Cross. The celestial equator is 0° of declination, which corresponds to celestial latitude. Right ascension corresponds to celestial longitude, with each hour of right ascension representing 15° of arc. Apparent brightness of stars is shown with dot sizes to illustrate the magnitude scale, where the brightest stars in the sky are 0th magnitude or brighter and the faintest stars that can be seen with the unaided eye at a dark site are 6th magnitude. (*Wil Tirion*)

The first three stars of the cross, clockwise from the brightest, are Acrux, Mimosa, and Gacrux.

Crux, whose stars were originally part of Centaurus but which were considered separately when they were used for navigation by sixteenth-century explorers, is pictured on the flags of Australia and New Zealand. The modern boundaries of the 88 constellations, including this one, were defined by the International Astronomical Union in 1928. *See* CENTAURUS; CONSTELLATION.

Jay M. Pasachoff

Cryobiology

The use of low-temperature environments in the study of living plants and animals. The principal effects of cold on living tissue are destruction of life and preservation of life at a reduced level of activity. Both of these effects are demonstrated in nature. Death by freezing is a relatively common occurrence in severe winter storms. Among cold-blooded animals winter weather usually results in a comalike sleep that may last for a considerable time. *See* HIBERNATION AND ESTIVATION.

While these natural occurrences suggest the desired ends of cryobiology, activities in this field differ in that much lower temperatures are employed than are present in natural environments. The extreme cold of liquid nitrogen (boiling at −320°F or −196°C) can cause living tissue to be destroyed in a matter of seconds or to be preserved for an indefinite time, certainly for years and possibly for centuries, with essentially no detectable biochemical activity.

The result achieved when heat is withdrawn from living tissue depends on processes occurring in the individual cells. Basic knowledge of the causes of cell death, especially during the process of freezing, and the discovery of methods which circumvent these causes have led to practical applications both for long-term storage of living cells or tissue (cryopreservation) and for calculated and selective destruction of tissue (cryosurgery). *See* CELL SENESCENCE AND DEATH.

Fundamental processes. The biochemical constituents of a cell are either dissolved or suspended in water. During the physical process of freezing, water tends to crystallize in pure form, while the dissolved or suspended materials concentrate in the remaining liquid. This principle has been suggested as one means for producing fresh water from seawater. In the living cell, however, this process is quite destructive. *See* WATER DESALINATION.

In a relatively slow freezing process ice first begins to form in the fluid surrounding the cells, and the concentration of dissolved materials in the remaining liquid increases. The equilibrium between the fluid inside the cell and the fluid surrounding the cell is upset. A concentration gradient is established across the cell wall, and water moves out of the cell in response to the osmotic force. As freezing continues, the cell becomes quite dehydrated. Salts may concentrate to extremely high levels. For example, sodium chloride may reach 200 g/liter (7 oz/qt) concentration in the final stages of freezing. In a similar manner the acid-base ratio of the solution may be altered during the concentration process. Dehydration can affect the gross organization of the cell and also the molecular relationships, some of which depend on the presence of water at particular sites. Cellular collapse resulting from loss of water may bring in contact intracellular components normally separated to prevent destructive interaction. Finally, as the ice crystals grow in size, the cell walls may be ruptured by the ice crystals themselves or by

the high concentration gradients which are imposed upon them.

By speeding the freezing process to the point that temperature drop is measured in degrees per second, some of these destructive events can be modified. For example, ice crystals will form within the cells so that less water will be lost, and the size of each crystal will be minimized, tending to reduce incidence of destruction of the cell walls. However, most of the previously mentioned destructive processes will prevail.

To prevent dehydration, steps must be taken to stop the separation of water in the form of pure ice so that all of the cell fluids can solidify together. The chief tools used to accomplish this are agents that lower the freezing point of the water. Typical materials used for industrial antifreeze agents are alcohols or glycols (polyalcohols). Glycerol, a polyalcohol which is compatible with other biochemical materials in living cells, is frequently used in cell preservation.

Besides the antifreeze additive, refrigeration procedures are designed to control the rate of decline in temperature to the freezing point, through the liquid-solid transition, and below, to very low temperatures. As a result of these procedures, the materials in the cell and in the surrounding fluid solidify together in a state more nearly resembling a glass than a crystalline material. In this form cells can be preserved apparently for as long as one may desire and will regain all their normal functions when properly thawed.

In the case of preservation of sizable aggregates of cells in the form of whole organs, the problems are significantly increased. With a large mass, control of the cooling rate is much more difficult. Additives may not reach cells inside a large organ in sufficient quantity to depress the freezing point adequately. The organ may suffer from oxygen deprivation or from other interruption of its normal environment quite apart from the freezing process itself. While there are signs of progress, whole-organ preservation is not yet a normally successful procedure.

Uses of cryopreservation. The earliest commercial application of cryopreservation was in the storage of animal sperm cells for use in artificial insemination. In the early post–World War II period artificial insemination of dairy herds was practiced widely. However, sperm were viable for only a few days when cooled but not frozen. The development of techniques for sperm freezing, storage, and transport using liquid nitrogen refrigeration, which occurred in the mid-1950s, greatly increased the use of artificial insemination, both for breeding stock and for human reproduction. In the frozen state sperm may be transported great distances. However, of even greater value is the fact that sperm cells can be stored for many years. *See* BREEDING (ANIMAL).

In an entirely similar manner the microorganisms used in cheese production can be frozen, stored, and transported without loss of lactic acid–producing activity. Pollen from various plants can be frozen for storage and transport. Normally this is of little value since fertile seeds can also be conveniently stored and transported without special preservation techniques. However, in plant-breeding experiments it is sometimes desired to crossbreed plants that do not grow in similar environments. In such instances the transport of frozen pollen can be utilized.

In biological research the growth of a pure strain or culture of a particular species is a necessity. For long-term studies these cultures must be continually renewed by growth on a suitable medium. Given enough time, these cultures undergo mutations; therefore, the species characteristics may change during the course of the investigation. By preparing a sufficiently large culture at the start of the investigation and freezing most of it, the investigator is assured of a continuous supply of unaltered microorganisms for his study no matter how long it may take. Banks of carefully established cell lines are stored in frozen condition for use by future generations of biologists.

Another application of cryopreservation is the storage of whole blood or separated blood cells. Frozen storage permits the buildup of the large stocks of blood required in a major catastrophe. It is also useful in storing the less common types of blood. Through frozen blood storage a person can actually develop a stock of their own blood for autotransfusion, a practice which is almost essential for those individuals with rare blood types. *See* BLOOD.

Freezing and preservation equipment. While other refrigerant materials or mechanical refrigeration systems can be used, most cryobiology apparatus has been designed to use liquid nitrogen as the refrigerant because it is convenient, biologically inert, relatively cheap, and plentiful. *See* NITROGEN.

When a controlled-rate freezer is used, the material to be frozen is placed in a well-insulated chamber. Liquid nitrogen is admitted to another portion of the

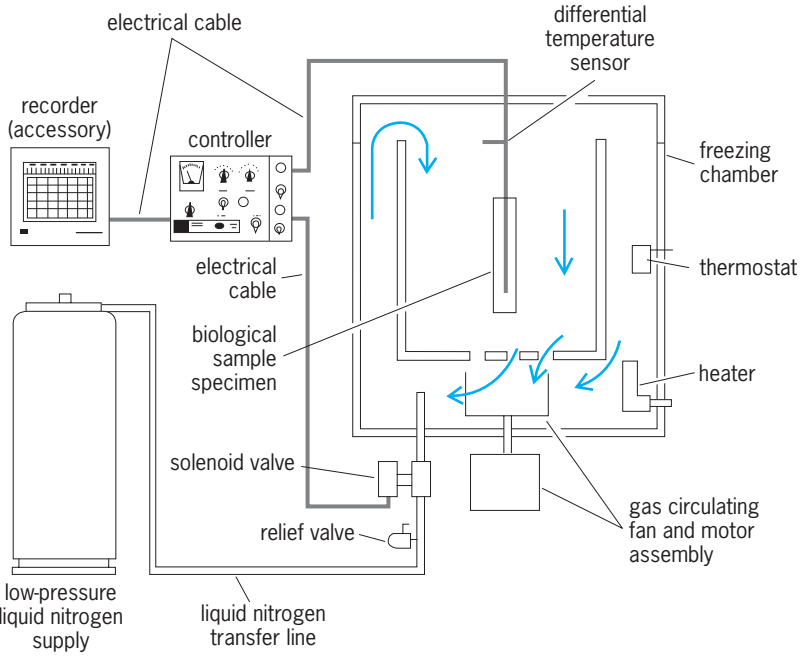
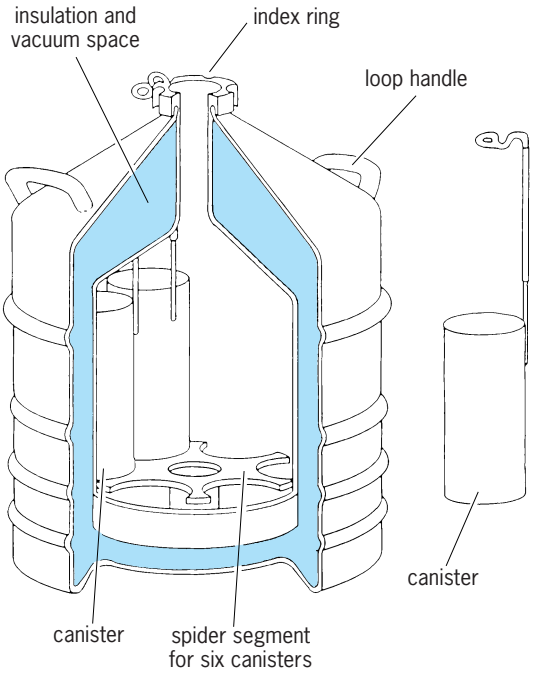


Fig. 1. Schematic representation of freezing system operation.

chamber at a rate controlled by an electrically operated valve. The liquid nitrogen is immediately vaporized by the air that is circulated through the chamber by a motor-driven fan. A thermocouple inserted in a sample of the material to be frozen is connected to the controller. Using the temperature of the thermocouple as a guide, the controller adjusts the flow of liquid nitrogen and the output of an electrical heater to obtain just the cooling rate desired. This procedure is outlined in **Fig. 1**. Working storage of sperm for artificial insemination or for small laboratories employs specially designed liquid nitrogen containers as illustrated in **Fig. 2**. A large opening (2.5-in. or 6.3-cm neck-tube diameter in diagram) is provided so that materials can be inserted and removed. Since the container is designed to allow very little leakage of heat into the inner container, one charge of liquid nitrogen can last for 60 days in active use. The frozen material is normally stored in small glass ampules. Six or eight ampules are clipped onto a vertical rod, and ten rods are mounted in a canister 2 in. (5 cm) in diameter. A typical storage container holds six canisters, or 360–480 ampules. A porous plug fits into the neck tube to hold the canister handles in place and to reduce the leakage of heat into the container.

While containers of the type just mentioned can be used to ship frozen biologicals by all normal means of transport, there are materials that would not conveniently fit into a container with such a small opening. Furthermore, some types of shipment require that the container be able to assume any position during transport. For these applications the equipment illustrated in **Fig. 3** has been developed. The entire top of this equipment is removable, leaving an opening as large as 14 in. (35 cm) in diameter. The liquid nitrogen refrigerant is absorbed in disks of porous material. The porous material can easily be shaped to conform to the contours of the object to be stored. Because the refrigerant liquid is absorbed in the porous packing disks, the container can be shipped in any attitude, although somewhat better results are obtained in the upright position.



Specifications:

height	= 23 3/8 in.
diameter	= 17 3/4 in.
neck tube diameter	= 2 1/2 in.
ampule capacity	= 360 (1.2-cc ampules, 6/rack)
	= 484 (1.0-cc ampules, 8/rack)

maximum liquid nitrogen capacity = 30 liters
working holding time = 60 days

Fig. 2. Liquid nitrogen container. 1 in. = 2.5 cm. 1 liter = 1.06 qt.

There is a container made for the stationary storage of large amounts of frozen material equipment. While this unit has much the same general appearance as a chest-type frozen food storage refrigerator, it is far superior in insulation performance. The unit can be left untended for a week without loss of effectiveness. The capacity of a large refrigerator is about 90,000 ampules (250 times the capacity of the portable container in Fig. 2), and several hundred units of frozen blood can be stored in it. Each item to be stored is mounted in a vertical rack. Stored material is retrieved by lifting the rack high enough to reach the desired package and then returning it to position. The brief exposure to ambient temperatures does not warm the other packages more than 1 or 2 degrees above their storage temperature of −320°F (−196°C).

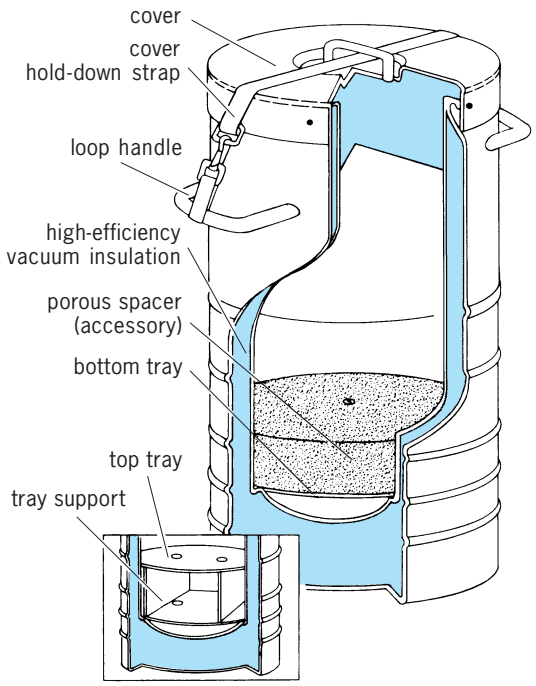
Cryosurgery. As was previously mentioned, the causes of cellular destruction from freezing can be effected deliberately to destroy tissue as a surgical procedure. A technique for controlled destruction of tissue in the basal ganglia of the brain was developed for relief from tremor and rigidity in Parkinson's disease.

One significant advantage of cryosurgery is that the apparatus can be employed to cool the tissue to the extent that the normal or the aberrant function is suppressed; yet at this stage the procedure can be reversed without permanent effect. When the surgeon is completely satisfied that the exact spot to be destroyed has been located, the temperature can be lowered enough to produce irreversible destruction. This procedure is of particular assistance in neurosurgery.

A second major advantage of cryosurgery is that the advancing front of reduced temperatures tends to cause the removal of blood and the constriction of blood vessels in the affected area. This means that little or no bleeding results from cryosurgical procedures.

A third major advantage of cryosurgery is that the equipment employs a freezing apparatus (which is about the size of a large knitting needle) that can be placed in contact with the area to be destroyed with a minimum incision to expose the affected area.

The equipment used in cryosurgery utilizes the same operating principles as the controlled-rate freezer. The flow of liquid nitrogen is controlled by an electrically operated valve whose function is directed by a controller that is based on the temperature sensed by a thermocouple located at the



Specifications:
height = 32 in.
diameter = 17 3/4 in.
neck tube diameter = 14 in.
ampule capacity = 5400 (1.2-cc ampules, 6/rack)
= 7200 (1.0-cc ampules, 8/rack)
working liquid
nitrogen capacity = 40.5 liters
working holding time = 19 days
bulk storage capacity = 1.6 ft
LD-40 used with accessory porous spacers
spacer thickness = 3 in.
spacer diameter = 13 1/2 in.
liquid nitrogen capacity
per spacer = 6.3 liters

Fig. 3. Liquid nitrogen container which is used for shipment of large samples. 1 in. = 2.5 cm. 1 ft = 0.3 m. 1 liter = 1.06 qt.

freezing element. Since it is desirable to freeze only the area immediately surrounding the tip of the instrument, the liquid nitrogen lead, the gaseous nitrogen return, the thermocouple leads, and the electrical supply for a small heater are all encased in a vacuum-insulated jacket.

Other uses. Besides the preservation or destruction of life, low temperatures can be employed for other purposes. For example, it may be desirable in the formulation of drugs to release antibodies or other active agents from their cells. One method by which this can be achieved is by grinding the cells while they are frozen. In frozen form the cells are hard and brittle and can be fractured easily without resulting in any chemical changes in the enzyme or antibody that is being released. A. W. Francis

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Cryogenics

The science and technology of phenomena and processes at low temperatures, defined arbitrarily as below 150 K (−190°F). Phenomena that occur at cryogenic temperatures include liquefaction and solidification of ambient gases (**Table 1**); loss of ductility and embrittlement of some structural materials such as carbon steel; increase in the thermal conductivity to a maximum value, followed by a decrease as the temperature is lowered further, of relatively pure metals, ionic compounds, and crystalline dielectrics (diamond, sapphire, solidified gases, and so forth); decrease in the thermal conductivity of metal alloys and plastics; decrease in the electrical resistance of relatively pure metals; decrease in the heat capacity of solids; decrease in thermal noise and disorder of matter; and appearance of quantum effects such as superconductivity and superfluidity. *See* ELECTRICAL RESISTIVITY; SPECIFIC HEAT OF SOLIDS; SUPERCONDUCTIVITY; SUPERFLUIDITY; THERMAL CONDUCTION IN SOLIDS.

Production of low temperatures. Cryogenic environments are maintained with cryogens (liquefied gases) or with cryogenic refrigerators. The temperature afforded by a cryogen (Table 1) ranges from its triple point to slightly below its critical point. Commonly used cryogens are liquid helium-4 (down to 1 K), liquid hydrogen, and liquid nitrogen. Less commonly used because of their expense are liquid helium-3 (down to 0.3 K) and neon. The pressure maintained over a particular cryogen controls its temperature. Heat input—both the thermal load and the heat leak due to imperfect insulation—boils away the cryogen, which must be replenished. *See* LIQUID HELIUM; THERMODYNAMIC PROCESSES; TRIPLE POINT.

A variety of techniques are available for prolonged refrigeration if such is favored by economic considerations or if very low temperatures are required. Down to about 1.5 K, refrigeration cycles involve

TABLE 1. Physical constants of low-temperature liquids

Liquid	Triple point		Normal boiling point, K	Critical point	
	Temp., K	Pressure atm†		Temp., K	Pressure, atm†
Helium-3			3.2	3.3	1.2
Helium-4			4.2	5.2	2.3
n-Hydrogen	14.0	0.07	20.4	33.2	13.0
n-Deuterium	18.7	0.17	23.7	38.4	16.4
n-Tritium	20.6	0.21	25.0	42.5*	19.4*
Neon	24.6	0.43	27.1	44.7	26.9
Fluorine	53.5	< 0.01	85.0	144.0	55.0
Oxygen	54.4	< 0.01	90.2	154.8	50.1
Nitrogen	63.2	0.12	77.4	126.1	33.5
Carbon monoxide	68.1	0.15	81.6	132.9	34.5
Argon	83.8	0.68	87.3	150.7	48.0
Propane	85.5	< 0.01	231.1	370.0	42.0
Propene	87.9	< 0.01	226.1	365.0	45.6
Ethane	89.9	< 0.01	184.5	305.4	48.2
Methane	90.7	0.12	111.7	191.1	45.8

*Calculated.
†1 atm = 1.013×10^5 pascals.

compression and expansion of appropriately chosen gases. At lower temperatures, liquid and solids serve as refrigerants. Adiabatic demagnetization of paramagnetic ions in solid salts is used in magnetic refrigerators to provide temperatures from around 4 K down to 0.003 K. Nuclear spin demagnetization of copper can achieve 5×10^{-8} K. Helium-3/helium-4 dilution refrigerators are frequently used for cooling at temperatures between 0.3 and 0.002 K, and adiabatic compression of helium-3 (Pomeranchuk cooling) can create temperatures down to 0.001 K. *See* ADIABATIC DEMAGNETIZATION.

Cryogenic refrigeration principles. In principle, any substance that undergoes an isentropic temperature change (meaning at constant entropy, without heat transfer or frictional-type losses) with a change in field intensity (pressure, stress, magnetic field intensity, or electric field intensity) can be used as a refrigerant. In practice, gases (for example, helium, hydrogen, and nitrogen) and magnetic substances (paramagnetic materials such as gadolinium sulfate or gadolinium gallium garnet) are usually employed.

An ideal refrigeration cycle, which has the maximum possible thermodynamic efficiency, is the Carnot cycle. This cycle can be illustrated with a temperature-entropy (*T*-*S*) diagram (**Fig. 1**), on which the state of the substance traces a rectangular path in a counterclockwise direction. Starting with the state of greatest temperature and entropy, the corners of the rectangle may be labeled *A, B, C*, and *D*. The *T*-*S* diagram is useful for understanding the thermodynamics of refrigeration cycles because the refrigeration (heat removed) is the area under the base of the rectangle (area *C*-*D*-*D*′-*C*′) and the net work input required is the area enclosed by the rectangle (area *A*-*B*-*C*-*D*). The magnetic substance (or gas) is magnetized (or compressed) isothermally from *A* to *B*; demagnetized (or expanded) isentropically from *B* to *C*, with cooling from temperature T_B to T_C; demagnetized (or expanded) further (*C* to *D*) at constant temperature with heat *Q* absorbed from the load given by Eq. (1); and finally magnetized (or compressed) back to the original state (*D* to *A*). The net work required to run the Carnot cycle is given by Eq. (2); therefore, the ratio of work input

$$Q = T_C(S_D - S_C) \tag{1}$$

$$W_C = (T_B - T_C) \cdot (S_D - S_C) \tag{2}$$

to refrigeration produced is given by Eq. (3). If T_A

$$\frac{W_C}{Q} = \frac{(T_B - T_C)}{T_C} \tag{3}$$

is the ambient temperature, Eq. (2) gives the ideal, or minimum, work required to supply refrigeration at temperature T_C. A useful, common measure of an actual cryogenic refrigerator's efficiency compares the ideal (Carnot) work to the actual work of refrigeration to give a so-called percent-of-Carnot, Eq. (4).

$$\eta_c = \frac{W_c}{W_{\text{actual}}} \tag{4}$$

See CARNOT CYCLE; ENTROPY; THERMODYNAMIC CYCLE.

Approximations to the Carnot cycle are actually used in magnetic refrigerators, with the upper temperature (T_B) ranging from 15 to 0.3 K, depending on the heat sink used (cryocooler, helium-4 bath, or helium-3 bath). Percent-of-Carnot efficiencies range up to about 50%. For gas refrigeration machines, however, the Carnot cycle is not practical because of heat transfer between the gas and the cylinder walls and the excessive pressures required to reach low temperature.

Two ideal thermodynamic cycles with practical application to gas refrigeration machines, however, are the reverse Stirling cycle and the reverse Brayton cycle with isothermal expander. These cycles may again be represented as counterclockwise paths on a temperature-entropy diagram (**Fig. 2**), labeled in the same way as before. Gas is isothermally compressed (*A*-*B*); cooled (*B*-*C*) in counterflow heat exchange by the return flow (*D*-*A*) in a thermal regenerator at constant volume (Stirling cycle) or counterflow heat exchanger at constant pressure (Brayton cycle); expanded isothermally (*C*-*D*), with the absorption of heat again given by Eq. (1); and then warmed at constant volume (Stirling cycle) or constant pressure (Brayton cycle) by counterflow heat exchange (*D*-*A*) with the incoming high-pressure (density) stream (*B*-*C*). The counterflow heat exchanger or regenerator is the fundamental element for attaining cryogenic temperatures with gas refrigerating machines, since it allows bootstrapping of the incoming flow to low temperature by heat exchange with the outgoing flow. Modest pressure ratios are possible because cooling by gas expansion is required for the thermal load (and irreversibilities in an actual machine) only, but not for initial attainment of the low temperature. *See* BRAYTON CYCLE; STIRLING ENGINE.

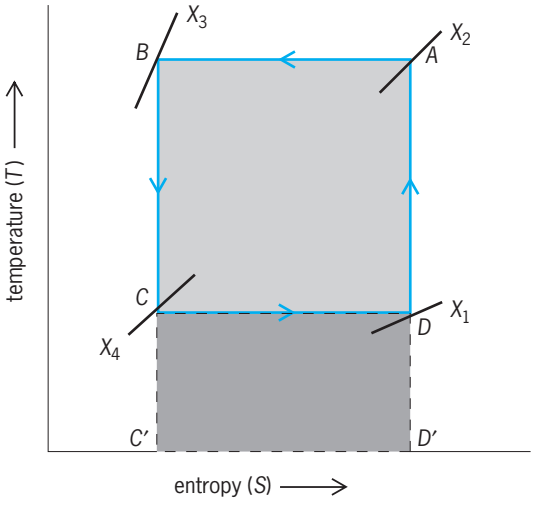

Fig. 1. Ideal Carnot refrigeration cycle on a temperature-entropy (*T*-*S*) diagram. The lines X_1, X_2, X_3, and X_4 represent lines of constant magnetization for a magnetic substance or constant pressure for a gas.

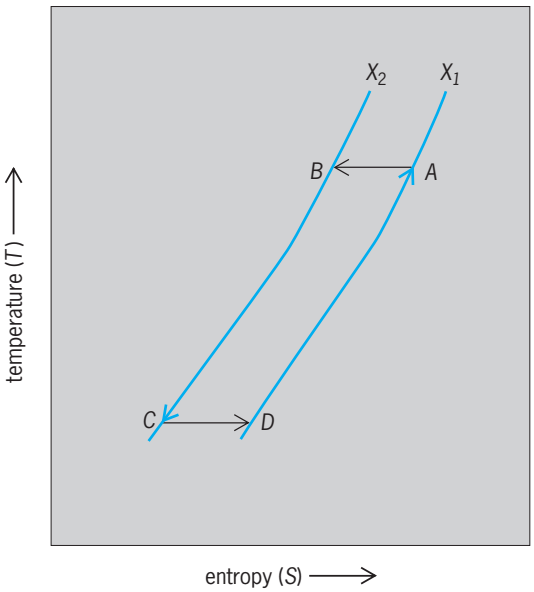

Fig. 2. Ideal practical gas refrigeration cycle on a temperature-entropy (*T*-*S*) diagram. The curves X_1 and X_2 are lines of constant volume in the reverse Stirling cycle and lines of constant pressure in the reverse Brayton cycle with isothermal expander.

Most modern cryogenic refrigerations are based on the reverse Stirling and reverse Brayton cycles. Reverse Stirling, Gifford-McMahon (G-M), and pulse-tube cryocoolers use thermal regenerators and expand the gas in the low-pressure part of the cycle with pistons or displacers (a gas piston for the pulse tube). Ordinarily, these machines are used for lower refrigeration loads (a fraction of a watt to a few hundred watts). Their lowest useful temperature is about 10 K because at lower temperatures the regenerator effectiveness becomes poor as a result of the rapidly diminishing heat capacity of the regenerator matrix. Experimental coolers reach as low as 4 K by using special high-heat-capacity regenerator materials. The reverse Brayton and the related Claude cycle use reciprocating piston expanders or, especially at higher capacities, turbo expanders. Temperatures as low as 1.8 K are typical for these machines. The percent-of-Carnot efficiency varies from as high as 40% for machines with refrigeration capacities in excess of 10 kW, to less than 1% for some fractional-watt machines. Higher refrigeration temperatures generally give slightly better efficiencies than lower refrigeration temperatures. Two refrigerators connected in series (for example, a magnetic refrigerator rejecting heat to a Stirling cooler) have an overall efficiency approximately equal to the product of the individual refrigerator efficiencies.

Gas liquefaction principles. Both the latent heat of vaporization and the sensible heat of the gas (heat content of the gas) must be removed to liquefy a gas. Of the total heat that must be removed to liquefy the gas, the latent heat is only 1.3% for helium and 46% for nitrogen. Consequently, an efficient liquefier must supply refrigeration over the entire temperature range between ambient and the liquefaction point, not just at the liquefaction temperature (**Table 2**).

The Collins-Claude refrigeration cycle (**Fig. 3**) forms the basis (with a multitude of variations) of most modern cryogenic liquefiers. Gas is compressed isothermally (1-2) and cooled (2-5) in a counterflow heat exchanger by the colder return stream of low-pressure gas (4′-1). During this cooling, a fraction of the high-pressure stream (equal to the rate of liquefaction) is split off (at points 2, 3, and 4) and cooled by the removal of work (energy) in expansion engines or turbines (which expand it to 2′, 3′, and 4′). This arrangement provides the cooling for the removal of the sensible heat. At the end of the counterflow cooling (point 5), the remaining high-pressure stream is expanded in either a Joule-Thomson valve or a wet expander to give the liquid product (point 6) and the return stream of saturated vapor (point 4′). *See* LIQUEFACTION OF GASES.

The work input required to produce refrigeration is commonly given in terms of watts of input power per watt of cooling, that is, W/W (Table 2). Cooling with a refrigerator is more efficient (that is, requires a lower W/W) than cooling with evaporating liquid supplied from a Dewar because the refrigerator does not discard the cooling available in the boil-off gas. *See* REFRIGERATION; REFRIGERATION CYCLE; THERMODYNAMIC CYCLE.

Storage and handling of cryogens. When gases are cooled from ambient to cryogenic temperatures, and liquefied, they undergo large increases in density, ranging from a 648-fold increase for methane to an 867-fold increase for hydrogen. Such an increase

TABLE 2. Ideal (reversible) power requirements for cryogenic liquefaction and refrigeration

Fluid	Temp., K	Liquefaction, W·h/liter*	Evaporating liquid refrigeration, W/W†	Ideal refrigerator, W/W‡
Helium	4.2	236	326	70.4
Hydrogen	20.4	278	31.7	13.7
Neon	27.1	447	15.5	10.1
Nitrogen	77.4	173	3.87	2.88
Fluorine	85	238	3.26	2.53
Argon	87.3	185	2.95	2.44
Oxygen	90.2	195	2.89	2.33
Methane	111.5	129	2.15	1.69

*Ideal (minimum) work required to change gas at ambient temperature and pressure to liquid.
†Ideal work to provide refrigeration at the cryogenic normal boiling point (column 2) by evaporating liquid without recovering the cooling available in the boil-off gas. This value is obtained by dividing the ideal work of liquefaction by the fluid's heat of vaporization.
‡Ideal work to provide refrigeration at the cryogenic normal boiling point with an ideal refrigerator such as Carnot or Stirling cycle machines.

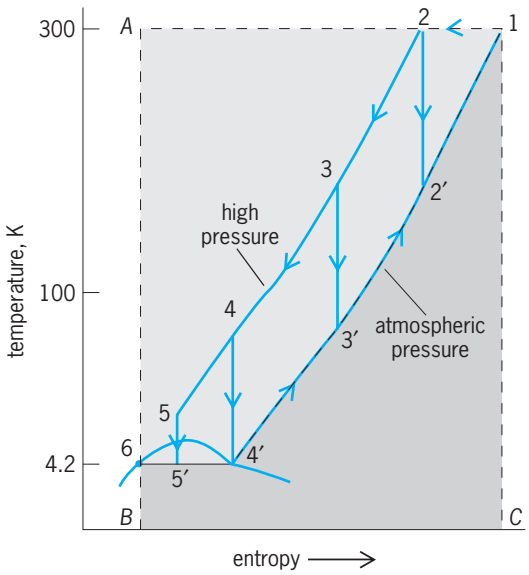

Fig. 3. Nearly ideal Collins-Claude gas liquefication cycle on a temperature-entropy (*T*-*S*) diagram. Area 1-4′-6-*B*-*C*−1 represents heat removed to liquefy the gas; area 1-*A*-6-4′−1 represents ideal (minimum) work required.

often makes it attractive to transport helium, hydrogen, oxygen, nitrogen, and natural gas as liquids, even when the end use requires a gas. The storage capacity of a trailer or rail car is about 10 times greater for liquid storage than gas storage.

Storage. Storage of cryogens requires special insulated containers. For helium and hydrogen, whose low heats of vaporization make them the most difficult to store, Dewars with evacuated multilayer insulation and boil-off gas-cooled radiation shields are used. Vacuums of 10^{-5} torr (10^{-3} pascal) or better virtually eliminate conduction heat ingress, while multilayer insulation, consisting of 100 or more layers of 0.25-mil (6-micrometer) aluminum-coated polyester film, and gas-cooled radiation shields greatly reduce radiation heat leak. Typical evaporative losses range from about 2% per day for 25-gallon (100-liter) liquid-helium Dewars to a fraction of a percent for large liquid-helium Dewars. For higher-boiling-point cryogens, with correspondingly higher heats of vaporization, such as liquid oxygen, liquid nitrogen, and liquefied natural gas (LNG), smaller containers use evacuated multilayer insulation, whereas larger vessels use evacuated perlite (silica powder) insulation or, for very large LNG tanks, unevacuated silica powder or foamed plastic. *See* DEWAR FLASK; HEAT INSULATION.

Transfer and metering. Transfer of cryogens requires pipes with insulations varying from evacuated multilayer insulation for liquid-helium service to foam insulation for large LNG lines. Transfer is by pressure, reciprocating pumps, or centrifugal pumps. For superfluid helium, the thermomechanical pump, which uses the unique ability of superfluid to flow without friction through porous plugs with micrometer size-channels, is also available. *See* PUMP.

The large economic value of cryogens, as well as the requirements of process control, make accurate metering important. Many standard meters such as orifice, venturi, turbine, positive-displacement, and coriolis are used, although two-phase flow and cavitation can reduce both accuracy and reliability. Liquid-level measurement in calibrated tanks is also used, especially for smaller vessels. *See* FLOW MEASUREMENT; LEVEL MEASUREMENT.

Applications. Cryogenic technology has manifold applications, including liquefaction of oxygen, nitrogen, and natural gas; transportation; aerospace; medicine; physics research; and electric power systems.

Oxygen and nitrogen. Every year the chemical, steel, and medical industries use millions of tons of oxygen and nitrogen. Although the end use is usually in the form of gas for breathing oxygen and applications such as basic oxygen steel making and production of ammonia and methanol, delivery is frequently in liquid form from cryogenic air-separation plants. *See* AIR SEPARATION; NITROGEN; OXYGEN.

Liquefied natural gas. Natural gas (methane with depending on the source, varying percentages of ethane, propane, butane, and nitrogen) is liquefied and shipped in insulated tankers with capacities up to 860,000 bbl (135,000 m^3). With the refrigeration power of some LNG plants exceeding 450 MW, these are the world's largest refrigeration systems. Storage vessels up to 1.6×10^6 bbl (250,000 m^3) receive LNG at the unloading sites.

Utility companies use LNG for either base load (continuous supply) or peak shaving (peak requirements). At some smaller peak-shaving plants, LNG is produced and stored during low-usage periods and returned as natural gas to the pipeline system during peak-usage periods. Some base-load installations contain air-separation facilities, so that refrigeration available from gasification of LNG can be applied to air liquefaction. *See* LIQUEFIED NATURAL GAS (LNG).

Transportation. LNG has considerable potential as a vehicle fuel because of its cost, efficient combustion, and especially low emissions. Compressed natural gas–powered vehicles are becoming commonplace. Liquid hydrogen has been demonstrated as both an aircraft and vehicle fuel, having the advantages of cleanliness and absence of carbon dioxide emission associated with the greenhouse effect. Prototype magnetically levitated trains, supported by superconducting magnets, are under evaluation. *See* AIRCRAFT FUEL; MAGNETIC LEVITATION.

Aerospace. As exemplified by the space shuttle, liquid hydrogen and liquid oxygen are the principal propellants for space flight because they have the highest specific impulse of any common fuel-oxidizer combination. Slush hydrogen, a mixture of liquid and solid hydrogen which coexist at 13.8 K (−434.8°F), has been studied extensively as a fuel for space probes and hypersonic aircraft-space vehicles. It offers a 16% greater density than normal-boiling-point hydrogen, as well as a greater capacity for cooling engine and aerodynamic surfaces. Supercritical hydrogen and oxygen are used on space vehicles to power fuel cells and to provide breathing oxygen.

Liquid oxygen, stowed in lightweight, well-insulated containers, is used in military aircraft for breathing oxygen and offers significant weight and space advantages over pressurized gas storage. *See* FUEL CELL; ROCKET PROPULSION.

Cryogenic wind tunnels, injecting liquid nitrogen to cool the flow, make possible matching of both Reynolds number and Mach number on aerodynamic models in the transonic region. Independent variation of the Reynolds and Mach numbers and lower power consumption are further advantages over conventional wind tunnels. *See* WIND TUNNEL.

Cooling infrared detectors in space satellites has been the impetus for extensive development of cryocoolers (small cryogenic refrigerators) which operate unattended for up to 5 years and supply refrigeration at 80 K (−316°F) and below. Considerable progress has been made, mostly with Stirling-cycle coolers, in attaining that required operating life. Superfluid helium at about 1.8 K is used as a coolant for infrared telescopes in astronomical satellites, and as a refrigerant in satellite experiments.

Medicine. Using liquid nitrogen, cryopreservation of human blood and bone and of both cattle and human sperm for artificial insemination is commonplace. Cryosurgery, involving destruction of tissue by freezing, is used to treat Parkinson's disease, cancers, and tumors (for example, pulmonary metastases, oral squamous cell carcinoma, malignant melanoma, and cancers of the brain, stomach and prostate). Routine cryotherapy removes warts without scarring. *See* CRYOBIOLOGY.

Magnetic resonance imaging (MRI) uses a large-bore superconducting magnet, cooled by liquid helium, to produce medical images of the body similar to x-ray tomography, but with better resolution of soft, primarily hydrogenic body tissue. Clinical use of these devices is widespread. *See* MEDICAL IMAGING.

Physics research. Liquid helium–cooled superconducting magnets, operating at 4 K (−452°F), are frequently used in high-energy particle accelerators for beam bending and focusing. Radio-frequency, cavities, cooled to about 2 K with superfluid helium, are used in electron-beam accelerators. Tokamak fusion reactors employ large, liquid helium–cooled magnets for confinement of the plasma. They are fueled by injecting solid deuterium pellets at speeds up to 1.2 mi/s (2 km/s) or more. These projects are quite large and employ some of the world's largest helium refrigerators and liquefiers. *See* NUCLEAR FUSION; PARTICLE ACCELERATOR.

Low-temperature physics focuses on superfluidity, superconductivity and, at temperatures below 1 millikelvin, the investigation of matter under conditions where thermal disorder is small. *See* LOW-TEMPERATURE PHYSICS.

Electric power systems. There have been a number of feasibility demonstrations of superconducting power systems cooled by liquid and supercritical helium. These include superconducting power transmission lines, superconducting generators and motors, low-loss superconducting transformers, and superconducting magnetic energy storage. *See* ENERGY STORAGE; SUPERCONDUCTING DEVICES.

Other applications. Additional applications of cryogenic technology include cryopumps for trapping gases on refrigerated surfaces or molecular sieves in vacuum systems, high-field-gradient magnetic separation and purification of materials, food preservation, hydrogen isotope separation using low-temperature distillation, metal treatment, and scrap processing. *See* FOOD PRESERVATION; ISOTOPE SEPARATION; VACUUM PUMP.

Superconductivity use. A number of the cryogenic applications discussed above use superconducting devices, especially magnets. Cooling these devices is a major area of cryogenic technology. The discovery of high-temperature superconductors, some with transition temperatures over 150 K (−190°F), shows promise of making economically feasible a number of applications that are currently uneconomical with low-temperature, liquid helium–cooled superconductors. *See* MAGNET.

D. E. Daney

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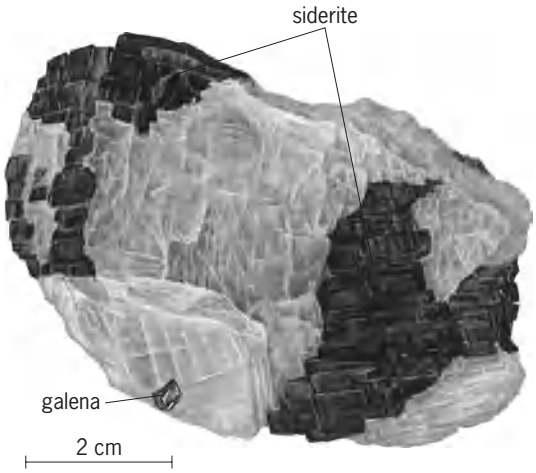
Cryolite

A mineral with chemical composition Na_3AlF_6. Although it crystallizes in the monoclinic system, cryolite has three directions of parting at nearly 90°, giving it a pseudocubic aspect. Hardness is $2^1/_2$ on Mohs scale and the specific gravity is 2.95. Crystals are usually snow-white but may be colorless and more rarely brownish, reddish, or even black (see **illus.**). The mean refraction index is 1.338, approximately that of water, and thus fragments become invisible



White translucent cryolite in association with siderite and galena, Ivigtut, Greenland. (*Specimen courtesy of Department of Geology, Bryn Mawr College*)

when immersed in water. Cryolite, associated with siderite, galena, and chalcopyrite, was discovered at Ivigtut, Greenland, in 1794. This locality remains the only important occurrence.

Cryolite was once used as a source of metallic sodium and aluminum, but now is used chiefly as a flux in the electrolytic process in the production of aluminum from bauxite. *See* ALUMINUM; BAUXITE.

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Cryotron

A current-controlled switching device based on superconductivity for use primarily in computer circuits. The early version has been superseded by the tunneling cryotron, which consists basically of a Josephson junction. In its simplest form (see **illus.**)

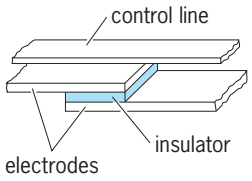

Diagram of the tunneling cryotron.

the device has two electrodes of a superconducting material (for example, lead) which are separated by an insulating film only about 10 atomic layers thick. For the electrodes to become superconducting, the device has to be cooled to a few degrees above absolute zero. The tunneling cryotron has two states, characterized by the presence or absence of an electrical resistance. They can be considered as the "on" and "off" states of the switch, respectively. Switching from on to off is accomplished by a magnetic field generated by sending a current through the control line on top of the junction. The device can switch in a few picoseconds and has a power consumption of only some microwatts. These properties make it an attractive switching device for computers, promising performance levels probably unattainable with other devices. *See* JOSEPHSON EFFECT; SUPERCONDUCTING DEVICES; SUPERCONDUCTIVITY.

P. Wolf

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Cryptobiosis

A state in which the metabolic rate of an organism is reduced to an imperceptible level. The several kinds of cryptobiosis (hidden life) include anhydrobiosis (life without water), cryobiosis (life at low temperatures), and anoxybiosis (life without oxygen). Of these, most is known about anhydrobiosis; therefore, the discussion will be restricted to that type.

States of anhydrobiosis occur only in the early developmental stages of various organisms, including plant seeds, bacterial and fungal spores, cysts of certain crustaceans, and certain insect larvae. They occur in both developmental and adult stages of certain soil-dwelling invertebrates (rotifers, tardigrades, and nematodes), mosses, lichens, and certain ferns.

Induction of anhydrobiosis. Many organisms that are capable of entering an anhydrobiotic state require a period of induction of the morphological and biochemical alterations. For example, rotifers, tardigrades, and nematodes include species that live in the water associated with soil. When this water dries up, the animals dry up also, but are not killed by the dehydration. When they come in contact with water again, they rapidly swell and assume active life, often within minutes. In order to enter this remarkable state the animals require slow dehydration, during which they undergo pronounced morphological changes. Tardigrades and rotifers contract longitudinally, assuming a so-called tun configuration (**Fig. 1**). Nematodes typically coil into tight spirals (**Fig. 2**). In all these animals, the morphological changes characteristic of cryptobiosis are thought to reduce the rate of evaporative water loss by reducing the animal's surface area accessible to the surrounding

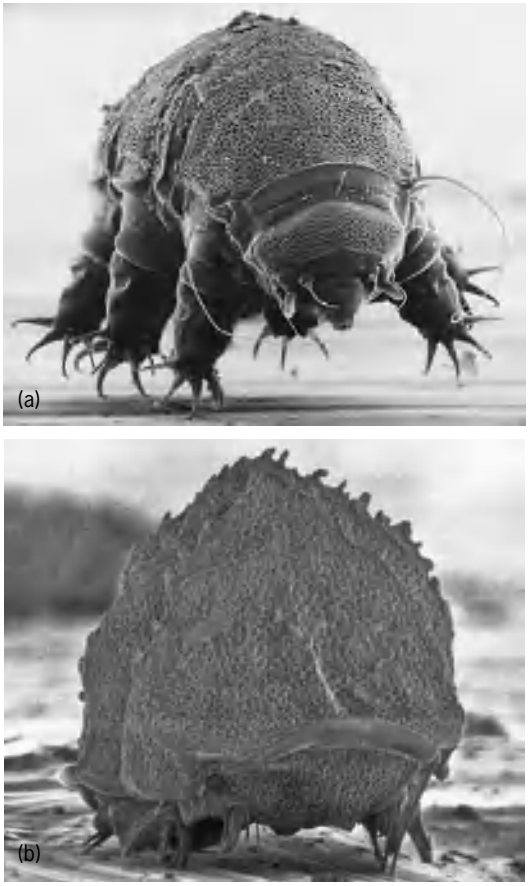

Fig. 1. Scanning electron micrographs of tardigrades: (*a*) active and (*b*) anhydrobiotic. (*From J. H. Crowe and A. F. Cooper, Jr., Cryptobiosis, Sci. Amer., 225:30–36, 1971*)

Fig. 2. Scanning electron micrograph of anhydrobiotic nematode. (*From J. H. Crowe and K. A. C Maden, Anhydrobiosis in nematodes: Evaporative water loss and survival, J. Exp. Zool., 193:323–334, 1975*)

air. In addition, contraction of the body is thought to entail ordered packing of organ systems and intracellular contents, thereby obviating mechanical damage.

Electron microscopy studies on dry organisms lend support to these suppositions. During the slow dehydration the animals show changes in their chemical composition. The nematode *Aphelenchus avenae*, for which there is the most information, synthesizes two carbohydrates, free glycerol and trehalose, that are thought to be important to its survival of dehydration. It appears that these molecules are made from storage lipids by way of the glyoxylate cycle, a biochemical pathway that is unusual among animals. Once the animals have produced these compounds, they can be dehydrated to less than 2% water content and can be kept in this dehydrated state for years. Workers in this field originally believed that both these compounds might be involved in stabilizing dry cells, but other studies have shown that while trehalose appears to be present at high concentrations in all animals that survive extreme dehydration, glycerol is not always present. Trehalose is also found at high concentrations in lower plants, such as yeast cells and the desert resurrection plant, that survive dehydration, but the analog of trehalose in higher plants appears to be sucrose. Thus, the data available from comparative biochemistry strongly suggest that the disaccharides trehalose and sucrose are involved in protecting dry organisms from damage due to dehydration.

In contrast with the above examples, cysts of crustaceans like the brine shrimp *Artemia* and seeds of plants seem to require no inductive period prior to anhydrobiosis. An *Artemia* cyst, for example, is an embryo covered with a capsule that is produced by the mother. The encapsulated embryo is released from the mother in a partially dehydrated state, and may afterward be rapidly reduced to low water content. Female *Artemia* also produce embryos without capsules that proceed with normal development without an intervening period of anhydrobiosis. In fact, these nonencapsulated embryos appear incapable of surviving dehydration. Encapsulated embryos contain large amounts of free glycerol and trehalose, similar to the amounts found in anhydrobiotic nematodes. Embryos produced without capsules contain insignificant amounts of glycerol and far less trehalose. Thus, induction of anhydrobiosis certainly occurs in *Artemia*, but the inductive mechanisms, which are poorly understood, probably act on the parent rather than directly on the embryo. Similar arguments can be made concerning the induction of anhydrobiosis in the seeds of higher plants and bacterial and fungal spores.

Biochemical adaptation and anhydrobiosis. Aside from gross mechanical stresses, dehydration results in a variety of other stresses that the chemical composition of anhydrobiotes may inhibit or obviate altogether. For example, it is well established that most macromolecules and cell membranes are maintained in a functional state by hydrophobic interactions. When water is removed, molecules and membranes lose their functionality; in proteins hydrophobic residues are exposed, and membranes undergo phase transitions that are likely to result in displacement of integral proteins from the membrane. A considerable body of evidence is accumulating which suggests that disaccharide compounds like trehalose may interact with macromolecules and membranes, stabilizing them at low water activities. For example, it has been possible to dry membranes in the presence of trehalose and to obtain membranes that show normal biological activity and structure upon rehydration. By contrast, when the same membranes are dried without trehalose, extensive structural damage occurs and all biological activity is lost. Similarly, proteins that are usually labile to drying can be dried in the presence of trehalose without loss of activity. *See* CELL MEMBRANES.

Another possible consequence of dehydration is that molecules, normally separated by bulk water in hydrated cells, would come into proximity, potentially resulting in chemical reactions between these molecules. Trehalose could serve as a substitute for bulk water, preventing such reactions. Yet another possible consequence of dehydration arises from the observation that oxygen is deleterious to anhydrobiotes. The organisms survive much longer in the dry state if oxygen is excluded.

It has been shown that damage to the dry organisms is probably due to direct oxidation of their chemical constituents, resulting in the formation of highly reactive free radicals; and that free glycerol inhibits such oxidation, which may explain the presence of glycerol in many dry but viable organisms. Finally, the reaction between reducing sugars and protein in the dry state, which is known as the browning reaction, could have serious consequences for dry tissues. This reaction involves the formation of a covalent bond between the protein and sugar, resulting in an insoluble product and denatured protein. It is possible that glycerol inhibits this reaction. It is not known whether sugars like trehalose show a browning reaction with dry proteins, since it is a nonreducing disaccharide of glucose.

Thus, it would be an advantage to the anhydrobiote to store low-molecular-weight carbohydrates in this form.

Longevity of anhydrobiotes. It is not known just how long anhydrobiotes can survive. Some nematodes have been revived after being kept dry for 39 years. A museum specimen of moss that had been kept dry for over 120 years yielded a number of viable rotifers and tardigrades. Cysts of *Artemia* survive under appropriate conditions (under vacuum and at temperatures below 32°F or 0°C) for at least 10 years and probably considerably longer. Possibly the most celebrated (and most debated) case of extreme longevity of anhydrobiotes is that of the viable lotus seeds recovered from a dry peat bed in Manchuria by Professor I. Ohga in 1927. These seeds were initially reported to be about 400 years old, but radiocarbon dating suggested the seeds were no more than 150 years old. Other, similar reports have appeared, but the greatest authenticated longevity for a seed was obtained from the accidental hydration and subsequent germination of a seed in the British Museum in the course of an air raid in 1942. The seed had been mounted on a herbarium sheet in 1702. Thus, though there is no definitive data on the limit of longevity of anhydrobiotes, it seems likely to be impressive—on the order of centuries.

A factor that may be related to the longevity of anhydrobiotes is their phenomenal resistance to environmental extremes. In the 1920s researchers found that anhydrobiotic rotifers and tardigrades could be exposed to temperatures as high as 304°F (151°C) and as low as −328°F (−200°C), provided their water content was low. It has also been shown that anhydrobiotic rotifers and tardigrades survived exposure to a temperature at the brink of absolute zero (0.008 K). The reported LD_{50} (median lethal dose) for x-ray exposure of anhydrobiotic tardigrades is 570,000 roentgens; by contrast, the LD_{50} for humans is about 500 roentgens. Tardigrades and nematodes will survive exposure to high vacuum (less than 10^{-6} mmHg or 10^{-4}Pa). In fact, the tardigrade shown in Fig. 1*b* and the nematode in Fig. 2 were each kept at such a vacuum and were bombarded with electrons for 15 min or longer in a scanning electron microscope. When the animals were moistened after removal from the microscope they both recovered, although the tardigrade died shortly afterward. Similar observations have been made on cysts of *Artemia*, seeds, and spores.

The resistance of anhydrobiotes to environmental extremes has important economic consequences. As a result, considerable research has been devoted to killing anhydrobiotes if they are detrimental to humans and their various agricultural endeavors or to preserving the anhydrobiotes if they are beneficial. For example, many soil-dwelling nematodes that exhibit anhydrobiosis are important agricultural pests. In the anhydrobiotic state they are remarkably resistant to chemical control measures. *Artemia* cysts and seeds, on the other hand, are beneficial to humans, and the fact that they can be stored and transported dry is particularly convenient. *Artemia* cysts have become vital links in aquaculture systems. The cysts are hydrated and hatched, and the young brine shrimp are fed to fish, crustaceans, and other aquatic animals maintained in culture.

Other applications of anhydrobiosis. The discovery that biological membranes and proteins can be dried in the presence of trehalose without loss of activity has led to applications that may be important in human welfare. For instance, liposomes are artificial membranes enclosing an aqueous compartment. These structures are being used for delivery of water-soluble drugs to cells. The liposomes are injected intravenously and are transported through the circulatory system. They fuse with cell membranes, depositing the drug in the cytoplasm of the cell. Ultimately, it may be possible to target specific cells with liposomes. One difficulty with this application is that liposomes are not stable structures. However, it is now possible to dry the liposomes with trehalose and to store them in this state. When the liposomes are needed, water can be added and they are ready for immediate use. There are similar applications in other areas of the pharmaceutical industry, medicine, and agriculture.

Metabolism of anhydrobiotes. A central question in the study of anhydrobiosis has been whether metabolism actually ceases. Most investigations in this area have involved attempts at measuring the metabolic rate of the dry organism directly. Increasingly sophisticated measurements have been made, using, for example, oil-filled cartesian divers to record oxygen uptake, radiotracers to study metabolism of radiolabeled metabolic intermediates, and gas chromatography to determine emission of carbon dioxide from the anhydrobiote. In all these studies the rate of metabolism recorded is low or undetectable. However, such observations cannot be construed to imply that metabolism is lacking altogether. The argument can always be made that metabolism does proceed in the anhydrobiote but the techniques used are not sufficiently sensitive to measure it. The most convincing argument is the one based on the hydration properties of proteins in laboratory cultures: If all the water contained by an anhydrobiotic organism were associated with protein, the protein would still exist in a semicrystalline state, and enzymes would not possess catalytic activity. Thus, enzyme-mediated metabolism must almost certainly cease at the low water contents that these organisms can possess. However, some chemical reactions do proceed in anhydrobiotes; mostly they are deleterious ones, such as the browning reaction. Such reactions could be construed as metabolism. But would it then follow that chemical reactions in rocks, air, and dead animals are metabolism, or even that a cosmic metabolism comprises the total chemical reactions of the universe.

Thus, the available evidence strongly suggests that dry anhydrobiotes are ametabolic. If that is the case, a philosophical question immediately arises concerning the nature of life. If metabolism is absent, the organism is generally referred to as dead. By this line of reasoning, anhydrobiotes would therefore be dead,

returning to life when they are rehydrated. But it is also known that some anhydrobiotes die while in the dry state, in the sense that an increasing proportion of the population fails to resume activity upon rehydration. It follows, then, that they must have "died" while they were "dead."

This philosophical quandary can be avoided by applying the definition of life adopted by most students of anhydrobiosis: An organism is alive, provided its structural integrity is maintained. When that integrity is violated, it is dead. *See* METABOLISM; PLANT METABOLISM.

John H. Crowe; Lois M. Crowe

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Cryptography

The various mechanisms for data confidentiality, data authentication, and nonrepudiation (that is, the property to prevent the denial of previous commitments or actions). Cryptography, with a long history of being an art of secret writing, has evolved into a science. As society is increasingly dependent upon computers, the vast amounts of data communicated, processed, and stored within computer systems and networks often have to be protected, and cryptography is a means of achieving this. It provides confidentiality services for transmitted data as well as stored data. Cryptographic techniques can also be used for message authentication, entity authentication, and digital signature verification for electronic funds transfer and credit card transactions. *See* COMPUTER SECURITY; DATA COMMUNICATION; DATABASE MANAGEMENT SYSTEM; DIGITAL COMPUTER; ELECTRICAL COMMUNICATIONS.

Basics. Data confidentiality is the property that only those authorized can access the data. Cryptography achieves data confidentiality by transforming meaningful messages, called plaintext, into unintelligible forms, called ciphertext or cryptogram. The process of transforming plaintext into ciphertext is called encryption or encipherment, and the opposite process is called decryption or decipherment. (An encryption scheme is also referred to as a cipher.) As shown in **Fig. 1**, the encryption algorithm is controlled by a parameter k', and the decryption algorithm is controlled by a secret k, known as a key. For confidentiality purposes, k should be kept

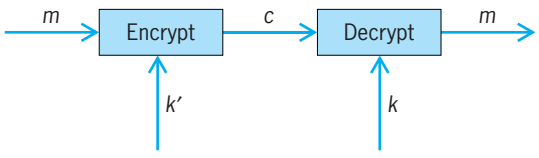

Fig. 1. The confidentiality model.

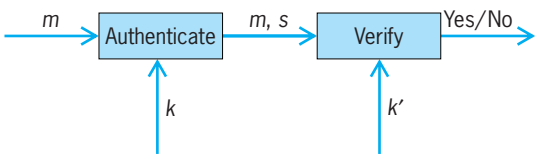

Fig. 2. The authentication model.

secret. If k' of a cipher can be used to derive k, then k' should also be kept secret, and such a cipher is called a symmetric key encryption. If it is computationally hard (see below) to derive k from k', then k' can be made public and the corresponding cipher is called an asymmetric encryption or a public key encryption. In most symmetric key encryption schemes, k' is the same as k and they are called symmetric keys. In public key cryptography, k' is called a public key and k is called a private key.

In a symmetric key encryption, if each encryption operation processes just one character or one bit, it is called a stream cipher. A block cipher, on the other hand, encrypts a block of fixed size, usually multiple bytes, each time.

In addition to data confidentiality, cryptography provides both data-origin authentication and entity authentication services. Data-origin authentication allows a message receiver to verify the message origin and its integrity. Entity authentication allows one to check the identity of another entity that possesses a certain secret. Given a message m, cryptography achieves data-origin authentication by appending to the message some additional bits (such as "s" in **Fig. 2**), called authentication tags, which are computed from m and a secret k. The receiver of the message and the authentication tags can verify the message origin and integrity by applying a verification parameter k' to the received data, including the authentication tags and the message. In most symmetric key cryptosystems, k and k' are the same and they should be kept secret. In public key cryptography, k' can be made public and k is kept secret; it is computationally hard (see below) to derive k from k'. Authentication tags generated by symmetric-key cryptography are called message authentication codes (MAC), which can be generated by both the sender and the receiver as both of them know k. In contrast, authentication tags by public key cryptography, called digital signatures, are stronger: since only the sender knows k and can produce a valid authentication tag under k, he cannot deny having sent the message in a later situation. Thus authentication tags by public key cryptography are the digital equivalent of a handwritten signature and have the potential for the nonrepudiation service.

Data confidentiality and data authentication are orthogonal in that one can exist without the other. On one hand, the authentication model given in Fig. 2 does not provide data confidentiality since message m is sent in the clear. On the other hand, encryption alone is insufficient to ensure that information is not altered during transmission. This fact is most evident when encryption with a public key algorithm is used. With a public key system, anyone can encipher

data by using the public enciphering key. Thus, any system user or node can masquerade as any other system user or node.

The opposite of cryptography is cryptanalysis, which studies techniques to defeat cryptographic methods. Cryptanalysts are those people doing cryptanalysis. The security of an encryption/authentication algorithm should depend solely on the secrecy of keys, not on the secrecy of algorithms. It is often assumed that the enemy cryptanalysts know the details of the algorithm employed.

In summary, symmetric block ciphers, symmetric stream ciphers, and public key encryption ciphers are used for data confidentiality, while MAC and digital signature algorithms are used for data authentication. We will first review the fundamental theory of these cryptographic algorithms and then give the details of some practical cryptographic algorithms.

Unbreakable algorithms. Unbreakable ciphers are possible, but the key must be randomly selected and used only once, and its length must be equal to or greater than that of the plaintext to be enciphered. Therefore such long keys, called one-time tapes, are not practical in data-processing applications.

Similarly, unconditionally secure authentication schemes (that is, schemes whose security does not depend on the computational power of the adversary) are possible. But the key must be randomly selected and used only once, which makes such keys impractical.

To work well, a key must be of fixed length, relatively short, and capable of being repeatedly used without compromising security. In theory, any algorithm that uses such a finite key can be analyzed; in practice, the effort and resources necessary to break the algorithm would be unjustified.

Strong algorithms. Fortunately, to achieve effective data security, construction of an unbreakable algorithm is not necessary. However, the work factor (a measure, under a given set of assumptions, of the requirements necessary for a specific analysis or attack against a cryptographic algorithm) required to break the algorithm must be sufficiently great. Included in the set of assumptions is the type of information expected to be available for cryptanalysis. For encryption schemes, this could be ciphertext only, plaintext (not chosen/known) and corresponding ciphertext, chosen plaintext and corresponding ciphertext, or chosen ciphertext and corresponding recovered plaintext.

A strong cryptographic algorithm must satisfy the following conditions: (1) The algorithm's mathematical complexity prevents, for all practical purposes, solution through analytical methods. (2) The cost or time necessary to unravel the message or key is too great when mathematically less complicated methods are used, either because too many computational steps are involved or because too much storage space is required.

To be strong, an encryption algorithm must satisfy the above conditions even when the analyst has the following advantages: (1) Relatively large amounts of plaintext (specified by the analyst, if so desired) and corresponding ciphertext are available. (2) Relatively large amounts of ciphertext (specified by the analyst, if so desired) and corresponding recovered plaintext are available. (3) Large high-speed computers are available for cryptanalysis.

Similarly, to be strong, an authentication algorithm must satisfy the above conditions even when the analyst has been given large amounts of messages (specified by the analyst, if so desired) and their corresponding authentication codes.

Computational complexity. The strength of a cryptographic scheme can be measured by the computational complexity of the task of cryptanalysis. The complexity of a task is the least possible number of elementary operations used by any program to accomplish this task. However, nontrivial lower bounds on the complexity of a problem are very hard to obtain. This is a fundamental problem in the design of cryptographic systems: it is very difficult to ensure that a system is sufficiently hard to crack. Without a good lower bound, the possibility that someone will find a fast algorithm for cryptanalyzing a given scheme must always be anticipated.

Problem classes P and NP. An important direction of theoretical work concerns the consideration (from computer science) of P versus NP. The class P consists of those problems which can be solved in polynomial time. That is, there are constants c and k such that, if the input to the problem can be specified in N bits, the problem can be solved on a sequential machine in time $c \times N^k$. Roughly speaking, these are the tractable problems. They include multiplication of two large numbers, exponentiation modulo a large prime, running the Data Encryption Standard (discussed below), and roughly any straightforward problem that does not involve searching.

The class NP (nondeterministic polynomial time) consists of problems which can be solved by searching. Roughly speaking, a possible solution to a problem in NP is to guess in turn each of 2^N possible values of some N-bit quantity, do some polynomial-time work related to each guess, and if some guess turns out to be correct, report the answer.

An example of a problem in NP is the knapsack problem: Given a set of integers $\{A_1, A_2, \ldots, A_n\}$ and a target integer B, can a subset of the A_i be selected without repetition (say $\{A_1, A_3, A_8\}$) such that their sum $(A_1 + A_3 + A_8)$ is the target B? One algorithm for solution is to try all possible subsets and just see whether any has the desired property. This algorithm requires exponential time, so called because the size of the input (n) occurs in the exponent of the formula expressing the running time (in this case, roughly 2^n). In fact, all known algorithms for this problem require exponential time. But there may be an unknown algorithm that runs in polynomial time; no proof prohibiting this is currently known.

Certainly, any problem in P is also in NP. A major outstanding question in computer science is whether P equals NP or whether NP is strictly larger.

There is a particular collection of problems in NP, including the knapsack problem, which are termed NP-complete. These can be thought of as the hardest

problems in NP. More precisely, an important mathematical result states: If there are any problems in NP but not in P, then each NP-complete problem is also in NP but not in P. This class has particular significance for cryptography.

In a good cryptographic system, certain operations (like decryption) are easy for those in possession of the key and difficult for those without the key. The cryptanalyst who could first guess and verify the correct key would be able to decrypt easily (in polynomial time). This could be done by searching over the space of possible keys and attempting to verify each one in polynomial time. Since the problem can be solved by searching, decryption is in NP.

If P = NP, then decryption would also be in P, and a good cryptographic system would most likely be difficult to design. If P ≠ NP, then the NP-complete problems might form a good starting point for cryptographic system design. Unfortunately, mathematicians and cryptographers have not yet learned how to transform an NP-complete problem into a secure cryptographic system.

Even if an NP-complete problem is eventually transformed successfully into a cryptographic system, a proof of the difficulty of cryptanalysis of that system (or any other) can be expected to be taxing. Such a proof would probably also prove that P ≠ NP, and this conjecture has eluded computer scientists so far. Thus, for now, the designers of cryptographic systems must rely on experience rather than rigorous proof of the difficulty of cryptanalysis.

Examples. Many practical and strong cryptographic algorithms have been developed. The Data Encryption Standard (DES), a block cipher, would require massive attacks to break it. The only known attacks are key exhaustion, which involves trying all $2^{56} \simeq 10^{17}$ possible keys, or attacks involving prohibitively large amounts (10^{14} bytes) of known plaintext and corresponding ciphertext. There is no guarantee against a more efficient, analytic attack, although none is known at present. A 56-bit key is not considered secure these days. A more advanced block cipher, Advanced Encryption Standard (AES), uses longer keys and larger data blocks and is assumed to be more secure and more efficient.

On the public key encryption vein, the RSA encryption algorithm, also discussed below, is based on the difficulty of factoring large numbers. However, a family of algorithms has been developed for factoring such numbers. If the modulus involved has, say, 2048 bits, then RSA should be secure in the foreseeable future. Other popular public key encryption algorithms include ElGamal public key encryption and Elliptic-curve (EC) public key encryption, both of which are based on the difficulty of the discrete logarithm problem.

Efficient MAC algorithms include those based on the cipher block chaining (CBC) mode of block ciphers, called CBC-MAC, and the MAC algorithms based on cryptographic hash algorithms, called HMAC.

Among the efficient digital signature algorithms are the RSA digital signature algorithm, the Digital Signature Standard (DSS, discussed below), and the Elliptic-curve Digital Signature Algorithm (ECDSA).

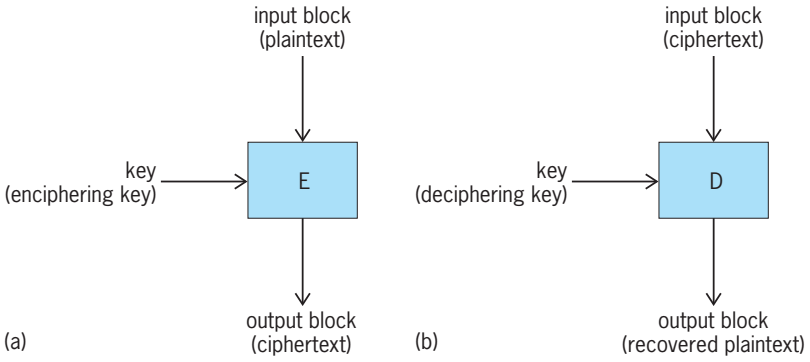

Fig. 3. Block cipher. (*a*) Enciphering. (*b*) Deciphering.

Block ciphers. A block cipher (**Fig. 3**) transforms a string of input bits of fixed length (called an input block) into a string of output bits of fixed length (called an output block). In a strong block cipher, the enciphering and deciphering functions are such that every bit in the output block jointly depends on every bit in the input block and on every bit in the key. This property is called intersymbol dependence.

Data encryption standard. Regardless of the application, a cryptographic system must be based on a cryptographic algorithm of validated strength if it is to be acceptable. The data encryption standard (DES) was such a validated conventional algorithm available in the public domain. It was accepted as a standard by the NIST in 1997 and was officially retired in 2004.

The DES enciphers a 64-bit block of plaintext into a 64-bit block of ciphertext under the control of a 56-bit key. Conceptually, it can be thought of as a huge key-controlled substitution box (S-box) with a 64-bit input and output. With such an S-box, $2^{64}!$ different transformations or functions from plaintext to ciphertext are possible. The 56-bit key used with DES thus limits the number of usable functions to 2^{56}.

A single, huge S-box is impossible to construct. Therefore DES is implemented by using several smaller S-boxes (with a 6-bit input and a 4-bit output) and permuting their concatenated outputs. By repeating the substitution and permutation process several times, cryptographic strength "builds up." The DES encryption process (**Fig. 4**) consists of 16 iterations, called rounds. At each round a cipher function (f) is used with a 48-bit key. The function is composed of the substitution and permutation. The 48-bit key, which is different for each round, is a subset of the bits in the externally supplied key.

The externally supplied key consists of 64 bits (56 bits are used by the algorithm, and up to 8 bits may be used for parity checking). By shifting the original 56-bit key, a different subset of 48 key bits is selected for use in each round. These key bits are labeled $K_1, K_2, \ldots, K_{16}$. To decipher, the keys are used in reverse order (K_{16} is used in round one, K_{15} in round two, and so on).

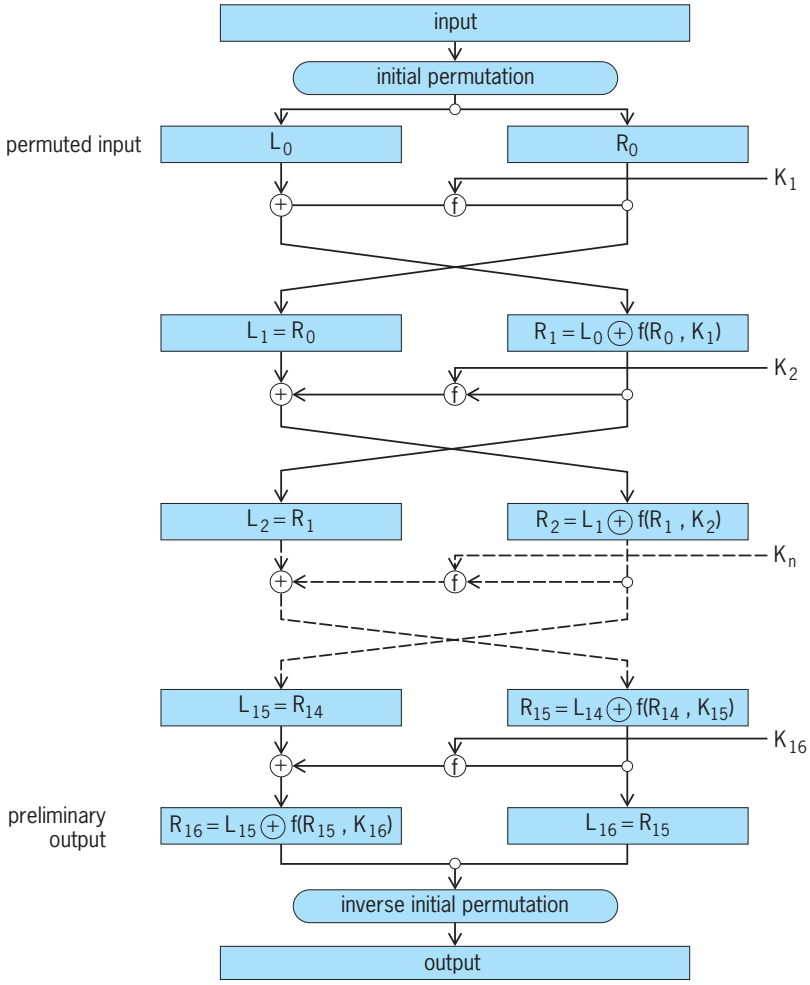

Fig. 4. Enciphering computation in the Data Encryption Standard.

At each round (either encipherment or decipherment), the input is split into a left half (designated L) and a right half (designated R) [Fig. 4]. R is transformed with f, and the result is combined, using modulo 2 addition (also called the EXCLUSIVE OR operation; $0 \oplus 1 = 1$, $1 \oplus 0 = 1$, $0 \oplus 0 = 0$, $1 \oplus 1 = 0$), with L. To consider one round of encipherment (**Fig. 5**), the input block (X) may be denoted $X = (L_0, R_0)$, where L_0 and R_0 are the left and right halves of X, respectively. Function f transforms R_0 into $f_{K1}(R_0)$ under control of cipher key K_1. L_0 is then added (modulo 2) to $f_{K1}(R_0)$ to obtain R_1, as in Eq. (1). The round is then completed by setting L_1 equal to R_0.

$$L_0 \oplus f_{K_1}(R_0) = R_1 \qquad (1)$$

The above steps are reversible, without introducing any new parameters or requiring that f be a one-to-one function. The ciphertext contains L_1, which equals R_0, and therefore half of the original plaintext is immediately recovered (**Fig. 6**). The remaining half, L_0, is recovered by recreating $f_{K1}(R_0)$ from $R_0 = L_1$ and adding it (modulo 2) to R_1, as in Eq. (2).

$$R_1 \oplus f_{K_1}(R_0) = [L_0 \oplus f_{K_1}(R_0)] \oplus f_{K_1}(R_0) = L_0 \qquad (2)$$

However, to use the above procedure for encipherment as well as decipherment, the left and right halves of the output must be interchanged; that is, the ciphertext (Y) is defined by Eq. (3). This modified procedure easily extends to n rounds, where the keys used for deciphering are $K_n, K_{n-1}, \ldots, K_1$.

$$Y = [L_0 \oplus f_{K_1}(R_0)], R_0 \qquad (3)$$

Advanced Encryption Standard (AES). The key length of DES is 56 bits, which is considered weak. Also, the 64-bit DES data block size is too small for modern applications. To overcome these difficulties, the National Institute of Standards and Technology (NIST) adopted an adapted version of the Rijndael algorithm, a symmetric block cipher, as the Advanced Encryption Standard (AES). AES uses 128-bit data block size and its keys can be 128, 192, and 256 bits, called *AES-128*, *AES-192*, and *AES-256*, respectively. Like DES, AES has multiple rounds: AES-128 has 10 rounds, AES-192 has 12 rounds, and AES-256 has 14 rounds. Different from DES, AES is not based on the Feistel cipher structure (Fig. 5).

In AES, the 16-byte plaintext is organized into a 4-by-4-byte array, called the State. The plaintext bytes are filled into the byte array in a column-wise way: the first column first, then the second column, the third column, and the last column. During the AES encryption, this state is continuously transformed and the end state is the ciphertext.

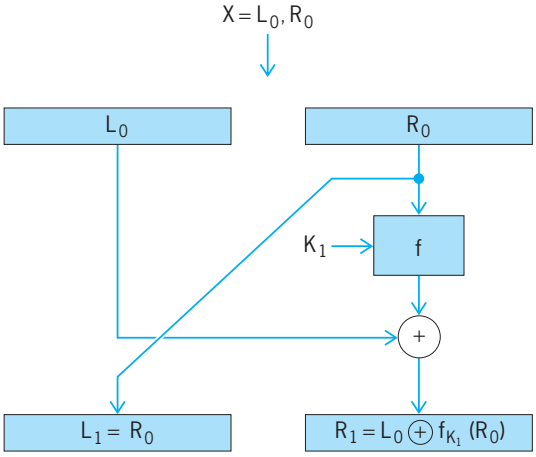

Fig. 5. Transformation of input block (L_0, R_0).

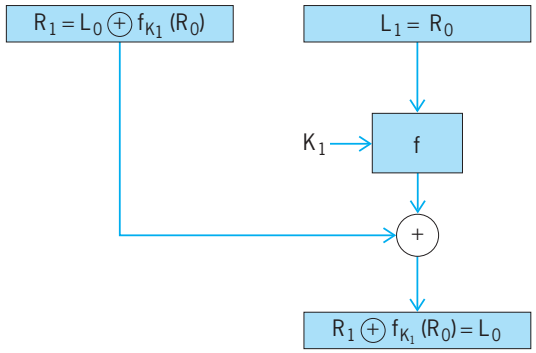

Fig. 6. Recovery of L_0.

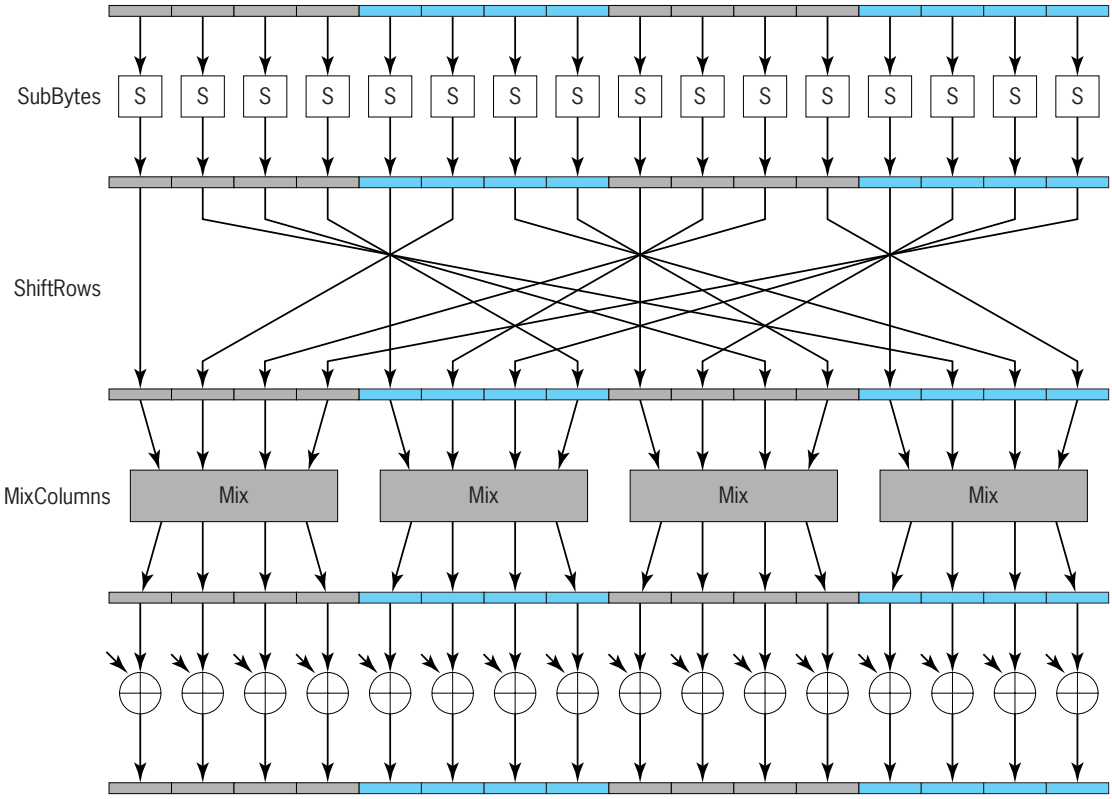

Fig. 7. AES round function.

For AES-128, AES-192, and AES-256, from the given cipher key, the AES key expansion operation derives, respectively, 11, 13, and 15 128-bit round keys.

The AES encryption starts with an initial round key addition, in which the first round key is added to the State by a simple bitwise XOR operation. Then, the State array is transformed by a round function multiple times (10 times for AES-128, 12 times for AES-192, and 14 times for AES-256), with the final round differing slightly with the other rounds. Each time the round function consumes one round key.

Figure 7 shows the details of the AES round function, which comprises four transformations, SubBytes, ShiftRows, MixColumns, and AddRoundKey. The SubBytes transformation is a nonlinear byte substitution (called S-box) and it works independently on a single byte. The ShiftRows transformation shifts around the bytes within the same rows of the State. The MixColumns transformation mixes the bytes within the same column of the State. The AddRoundKey transformation applies the corresponding round key to the State by a simple bitwise XOR operation. In the last AES round, the MixColumns transformation does not appear.

All the four AES transformation operations are reversible, which makes AES decryption possible. More specifically, AddRoundKey is its own inverse; the inverse of SubBytes is InvSubBytes, in which an inverse S-box is applied to the each byte of the State; the inverse of the ShiftRows transformation is InvShiftRows; and the inverse of the MixColumns transformation is InvMixColumns. The AES decryption is performed through a sequence of the inverse transformations, in reverse order, with the corresponding round keys generated from the same key expansion operation. This makes the AES decryption structure completely different from the AES encryption structure. AES decryption can also be structured in a similar way as the AES encryption, but a different key expansion operation should be employed.

Modes of operation. As a rule, block encryption is used to protect keys. A different method, called chained block encryption, is used to protect data. In chaining, the process of enciphering and deciphering is made dependent on other (prior) data, plaintext, ciphertext, and the like, also available at the time enciphering and deciphering take place. Thus every bit in a given output block depends not only on every bit in its respective input block and every bit in the key, but also on any or all prior data bits, either inputted to or produced during the enciphering or deciphering process.

Sometimes data to be enciphered contain patterns that extend beyond the cipher's block size. These patterns in the plaintext can then result in similar patterns in the ciphertext, which would indicate to a cryptanalyst something about the nature of the plaintext. Thus, chaining is useful because it significantly reduces the presence of repetitive patterns in the ciphertext. This is because two equal input blocks encipher into unequal output blocks.

A recommended technique for block chaining, referred to as cipher block chaining (CBC), uses a

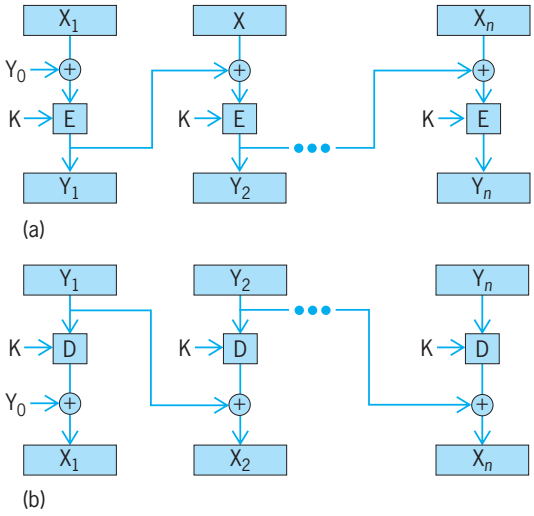

Fig. 8. Block chaining with ciphertext feedback. (*a*) Enciphering. (*b*) Deciphering.

ciphertext feedback (**Fig. 8**). Let $X_1, X_2, \ldots, X_n$ denote blocks of plaintext to be chained using key K; let Y_0 be a nonsecret quantity defined as the initializing vector; and let $Y_1, Y_2, \ldots, Y_n$ denote the blocks of ciphertext produced. The ith block of ciphertext (Y_i) is produced by EXCLUSIVE ORing Y_{i-1} with X_i and enciphering the result with K, as in Eq. (4) , where $\oplus$ denotes the EXCLUSIVE OR operation, or modulo 2

$$E_K(X_i \oplus Y_{i-1}) = Y_i \quad i \geq 1 \tag{4}$$

addition. Since every bit in Y_i depends on every bit in X_1 through X_i, patterns in the plaintext are not reflected in the ciphertext.

The ith block of plaintext (X_i) is recovered by deciphering Y_i with K and EXCLUSIVE OR-ing the result with Y_{i-1}, as in Eq. (5). Since the recovered

$$D_K(Y_i) \oplus Y_{i-1} = X_1 \quad i \geq 1 \tag{5}$$

plaintext X_i depends only on Y_i and Y_{i-1}, an error occurring in ciphertext Y_j affects only two blocks of recovered plaintext (X_j and X_{j+1}).

Other chained block encryption methods include cipher feedback (CFB), output feedback (OFB), and counter mode (CTR).

Stream ciphers. A stream cipher (**Fig. 9**) employs a bit-stream generator to produce a stream of binary digits R (traditionally termed the key stream), which is then combined either with plaintext X to produce ciphertext or with ciphertext Y to recover plaintext. In most stream ciphers, an EXCLUSIVE OR operation, or modulo 2 addition, is used to combine the respective bit streams. Thus encipherment and decipherment are defined by $X \oplus R = Y$ and $Y \oplus R = X$, respectively.

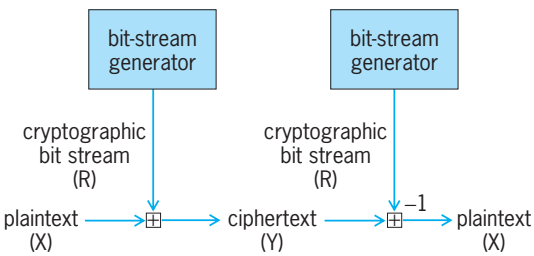

Fig. 9. Stream cipher concept.

If the bit-stream generator were truly random, an unbreakable cipher could be obtained by EXCLUSIVE ORing the plaintext and the cryptographic bit stream R. The cryptographic bit stream would be used directly as the key, and would be equal in length to the message. But in that case the cryptographic bit stream must be provided in advance to the communicants via some independent and secure channel. This introduces insurmountable logistic problems for heavy data traffic. Hence, for practical reasons, the bit-stream generator must be implemented as an algorithmic procedure. Then both communicants can generate the same cryptographic bit stream—provided that their algorithms are identically initialized (**Fig. 10**).

When modulo 2 addition is used as the combining operation, each bit in the output ciphertext (recovered plaintext) is dependent only upon its corresponding bit in the input plaintext (ciphertext). This is in marked contrast to the block cipher. Both approaches, however, have comparable strength.

For security purposes, a stream cipher must never predictably start from the same initial condition, thereby producing the same cryptographic bit stream. This can be avoided by making the cryptographic bit stream dependent on a nonsecret quantity Z (known as seed, initializing vector, or fill), which is used as an input parameter to the ciphering algorithm (Fig. 9). *See* INFORMATION THEORY.

The most famous symmetric stream cipher might be RC4.

RSA public key algorithm. The RSA algorithm (named for the algorithm's inventors, R. L. Rivest, A. Shamir, and L. Adleman) is a public key encryption algorithm based on the fact that factoring large composite numbers into their prime factors involves an overwhelming amount of computation.

To describe the RSA algorithm, the following quantities are defined:

p and q are primes	(secret)
$N = p \cdot q$	(nonsecret)
$\varphi(N) = (p-1)(q-1)$	(secret)
d is the private key	(secret)
e is the public key	(nonsecret)
m is the plaintext	(secret)
c is the ciphertext	(nonsecret)

Based on an extension of Euler's theorem (6) the

$$m^{x\varphi(N)+1} \equiv m \bmod N \tag{6}$$

algorithm's public and private keys (e and d) are chosen so that Eq. (7) or, equivalently, Eq. (8) [which is

$$d \times e = x\varphi(N) + 1 \tag{7}$$

$$d \times e = 1 \bmod \varphi(N) \tag{8}$$

used here in place of a more complicated equation for ease of understanding] is satisfied. By selecting two secret prime numbers p and q, the user can calculate $N = p \times q$, which is made public, and $\varphi(N) = (p - 1) \times (q - 1)$, which remains secret and is used to solve Eq. (8). (Tests are available to determine with a high level of confidence if a number is prime or not.) To obtain a unique solution for the public key (e), a random number, or secret key (d), is selected that is relatively prime to $\varphi(N)$. (Integers a and b are relatively prime if their greatest common divisor is 1.) e is the multiplicative inverse of d, modulo $\varphi(N)$, and e can be calculated from d and $\varphi(N)$ by using Euclid's algorithm. Equation (6) can therefore be rewritten as Eq. (9), which holds true for any plain-

$$m^{d \times e} \equiv m \bmod N \tag{9}$$

text (m). Encipherment and decipherment can now be interpreted as in Eqs. (10) and (11). Moreover, be-

$$E_e(m) = c \equiv m^e \bmod N \tag{10}$$

$$D_d(c) \equiv c^d \bmod N \equiv m^{e \times d} \bmod N \equiv m \bmod N \tag{11}$$

cause multiplication is a commutative operation ($e \times d = d \times e$), encipherment followed by decipherment is the same as decipherment followed by encipherment. Thus the RSA algorithm can be used for both confidentiality and digital signatures (see below).

Finally, since $m^e (\bmod N) \equiv (m + xN)^e (\bmod N)$ for any integer x, $E_e(m) = E_e(m + xN)$. Thus the transformation from plaintext to ciphertext, which is many-to-one, is made one-to-one by restricting m to the set $\{0, 1, \ldots, N - 1\}$.

Message authentication codes. Message authentication codes are based on symmetric key cryptography, and they assume that the message sender and receiver share a common symmetric key.

The strength of a MAC algorithm depends on its output length, its key length, and the algorithm itself. The strongest attack against a MAC algorithm is the key recovery attack, in which an attacker can recover the symmetric key from some messages and their corresponding MACs under that key. A weaker attack is the forgery attack: an attacker might not be able to find the symmetric key used, but can still forge a valid MAC on a given new message, using help from some known/chosen messages and their corresponding MACs under a certain key. A secure MAC algorithm should be secure against both types of attacks.

Two widely used MAC algorithms are the CBC-MAC, a MAC algorithm based on CBC mode of block ciphers, and HMAC, a MAC algorithm based on cryptographic hash algorithms such as SHA-1. In a CBC-MAC, the symmetric key is used as a block cipher key to encrypt the given message in CBC mode; the last block ciphertext is the MAC code (other ciphertext blocks are thrown away).

In a HMAC, a cryptographic hash algorithm, h, is used. Given a symmetric key k and a message m, its HMAC code is computed as $h(k \oplus \text{opad} \parallel h(k \oplus \text{ipad} \parallel m))$, where opad and ipad are specified constants.

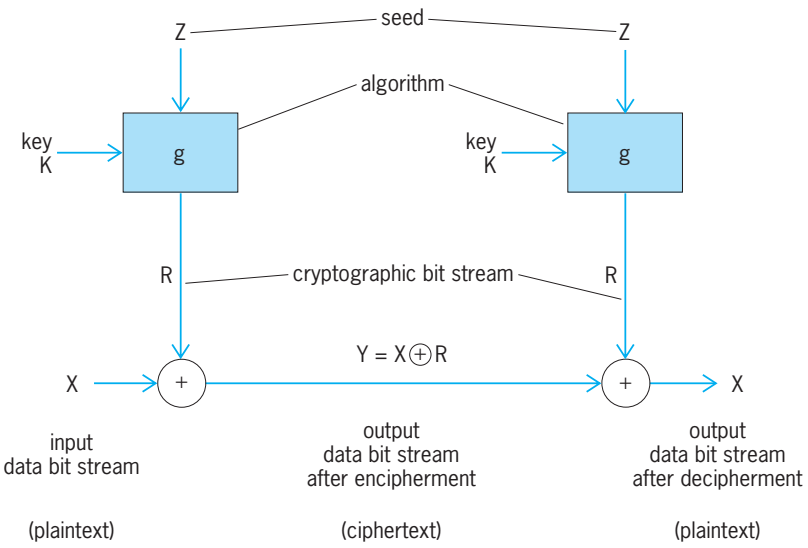

Fig. 10. Stream cipher using an algorithmic bit stream generator, modulo 2 addition, and seed.

Digital signatures. Digital signatures authenticate messages by ensuring that the sender cannot later disavow messages, the receiver cannot forge messages or signatures, and the receiver can prove to others that the contents of a message are genuine and that the message originated with that particular sender. The digital signature is a function of the message, a secret key or keys possessed by the sender and sometimes data that are nonsecret or that may become nonsecret as part of the procedure (such as a secret key that is later made public).

Digital signatures are more easily obtained with public key than with conventional algorithms. When a message is enciphered with a private key (known only to the originator), anyone deciphering the message with the public key can identify the originator. The originator cannot later deny having sent the message. Receivers cannot forge messages and signatures, since they do not possess the private key.

Since enciphering and deciphering keys are identical in a conventional algorithm, digital signatures must be obtained in some other manner. One method is to use a set of keys to produce the signature. Some of the keys are made known to the receiver to permit signature verification, and the rest of the keys are retained by the originator in order to prevent forgery.

The RSA algorithm previously discussed can be used for digital signatures. The National Institute of Standards and Technology (NIST) proposed another method for generating digital signatures, based on a 1985 proposal by T. El Gamal. Its security relies on the difficulty of computing logarithms over finite fields, or discrete logarithms, that is, of solving the following problem: Given a prime p and integers g, y, find an integer x satisfying $g^x \equiv y \,(\bmod p)$. In this scheme, all users share a large prime number p and a generator g (that is, an integer whose powers give all nonzero residues modulo p). Each user A has a

secret key x_A and a public key $y_A \equiv g^{(x_A)} \pmod{p}$. To each message m is associated an integer $h(m)$, $0 < h(m) < p$. User A "signs" the message by producing a pair of integers (r, s) between 0 and p satisfying Eq. (12). [The signature is computed by choosing a

$$g^{h(m)} \equiv y_A^r r^s \tag{12}$$

random integer k between 0 and $p - 1$, relatively prime to $p - 1$, computing $r \equiv g^k \pmod{p}$, and solving for s in Eq. (13): different random choices k will

$$h(m) \equiv x_A r + ks \bmod (p - 1) \tag{13}$$

give different signatures.] This "signature" is verified by using modular exponentiation of large integers. But a forger, without knowledge of the secret quantity x_A, apparently cannot produce such a signature without first solving the difficult discrete logarithm problem. *See* NUMBER THEORY.

A signature in the ElGamal scheme consists of two integers, (r, s), each about as large as the prime p. Since p must have at least 512 bits to make the discrete logarithm problem infeasible, the signature would then require 1024 bits.

A modification incorporated into the NIST proposal allows for shorter signatures. Here p is chosen such that $p - 1$ has a moderate prime factor $q \simeq 2^{160}$, and g is chosen so that its powers mod p form a subgroup of size q, so that s can be specified with fewer bits (160). The discrete logarithm problem is now restricted to that subgroup, which may or may not make cryptanalysis easier.

A caution in the design of digital signature schemes is the so-called birthday paradox: if signatures are n bits long, then among a collection of $(2^n)^{1/2} = 2^{n/2}$ messages there are likely to be two with the same signature. An opponent who finds it feasible to generate $2^{n/2}$ messages and their signatures can use the two that match to generate a signature. To thwart an opponent whose computing power might be 2^{64}, a digital signature of at least 128 bits is needed. The ElGamal, NIST, and RSA schemes need larger signatures than that because of algebraic shortcuts.

Xunhua Wang; Don Coppersmith;
Stephen M. Matyas; Carl H. Meyer

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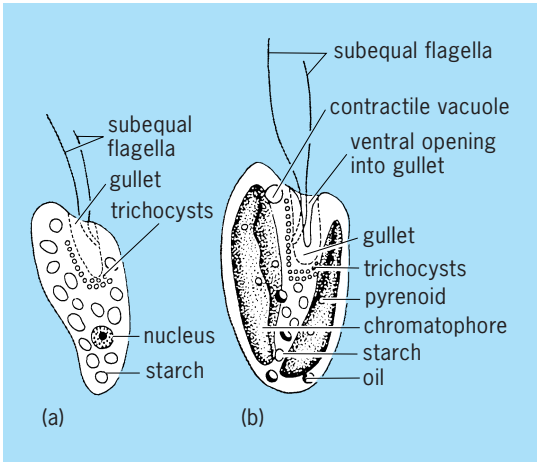
Cryptomonadida

An order of the class Phytamastigophorea, also known as the Cryptomonadina. Cryptomonads are considered to be protozoa by zoologists and algae by botanists. Most species occur in fresh water. They are olive-green, blue, red, brown, or colorless. All flagellated members have two subequal flagella inserted in a gullet opening through an obliquely truncate anterior end (see **illus.**). In *Nephroselmis* and *Protochrysis* the gullet is laterally displaced and a contractile vacuole opens into it. The nucleus is posterior. Oil and starch or a related compound are formed and pyrenoidlike bodies are present. *Cyathomonas* and *Chilomonas* are colorless; the latter has been widely used in nutritional studies, and extensive studies have been made of its cytology, demonstrating trichocysts around the gullet and muciferous bodies under the pellicle.



Examples of cryptomonads. (*a*) *Chilomonas paramecium*. (*b*) *Cryptomonas erosa*.

Phaeoplax is a palmelloid type with nonmotile cells embedded in a gelatinous matrix. The reproductive cells are flagellated. Radiolarians have symbiotic yellow cells, zooxanthellae, which are cryptomonads. Although few in genera and species, Cryptomonadida form extensive marine and freshwater blooms. *See* CILIA AND FLAGELLA; PHYTAMASTIGOPHOREA; PROTOZOA.

James B. Lackey

Cryptophyceae

A small class of biflagellate unicellular algae (cryptomonads) in the chlorophyll *a–c* phyletic line (Chromophycota). In protozoological classification, these organisms constitute an order, Cryptomonadida, of the class Phytomastigophora. Cryptomonads are 4–80 micrometers long, ovoid or bean-shaped, and dorsoventrally flattened. The cell is bounded by a moderately flexible periplast comprising the plasmalemma and underlying rectangular or polygonal proteinaceous plates. A tubular invagination (gullet, groove, or furrow) traverses the ventral cytoplasm and opens just below the apex of the cell. A pair of subequal flagella, which are covered with hairs and small scales, arise from the center or apical end of the gullet.

Cryptomonads may be photosynthetic, osmotrophic, or phagotrophic. In photosynthetic species,

chloroplasts occur singly or in pairs. The cryptomonad that is symbiotic in the ciliate *Mesodinium* is exceptional in having numerous chloroplasts. The photosynthetic lamellae, which are composed of loosely appressed pairs of thylakoids, are surrounded by four membranes. As in other members of the Chromophycota, the inner pair constitutes the chloroplast envelope while the outer pair represents endoplasmic reticulum confluent with the outer membrane of the nuclear envelope. The periplastidial compartment (space between the pairs of membranes on one side of the chloroplast) is more extensive than in other chromophytes. In other algae the compartment contains a dense network of tubules and vesicles, but in cryptomonads it contains starch grains, ribosomes, and a unique structure called a nucleomorph, in addition to scattered tubules and vesicles. The nucleomorph, which is smaller than the chloroplast, is bounded by a double membrane with numerous pores. It contains fibrillogranular inclusions. One nucleomorph, which always divides before the cell enters mitotic prophase, is closely associated with each chloroplast. If it can be shown that the nucleomorph is a vestigial nucleus, as it seems to be, the hypothesis that cryptomonad chloroplasts are evolutionarily derived from an endosymbiont will be supported.

Photosynthetic pigments include chlorophyll *a* and *c*, α-carotene and β-carotene (the former being predominant, an unusual ratio for algae), and red and blue phycobiliproteins closely related to those in red algae but occurring in the intrathylakoidal space rather than forming phycobilisomes. The color of the chloroplasts, and thus the color of the cryptomonad, depends upon pigment composition and may be green, olive, brown, yellow, red, or blue. Pigment composition is affected by environmental conditions. Thus, red cryptomonads at a low nitrate level gradually and reversibly bleach to green. Pyrenoids, usually capped with starch, are present in the chloroplasts of most species while eyespots are present in a few.

A single wormlike or reticulate mitochondrion, differing from those of other chromophytes in having flattened rather than tubular or inflated cristae, is present in each cell. Most cryptomonads have two sizes of ejectile organelles (trichocysts or ejectosomes)—those large enough to be seen with the compound microscope form rows along the gullet, and those much smaller are situated at corners of the periplast plates. Discharge of trichocysts imparts a characteristic darting movement to the cell. Contractile vacuoles may be present in fresh-water species.

Reproduction is by longitudinal binary fission. Sexual reproduction has not been confirmed.

Cryptomonads are found in fresh, brackish, marine, and hypersaline bodies of water, sometimes in great abundance. Several species are endosymbionts of marine ciliates and invertebrates, and of fresh-water dinoflagellates. They do not appear to be closely related to any other group of algae and thus are sometimes assigned to their own division. In addition to typical cryptomonads (order Cryptomonadales), of which about 200 species have been described, there are two monospecific genera with a life history that includes only a brief motile phase. In these genera, which constitute the order Tetragonidiales, the cells usually are nonflagellate and sedentary and are said to have a cellulose wall. They form zoospores, however, which are typically cryptomonad in structure. *See* ALGAE; CHROMOPHYCOTA; CRYPTOMONADIDA. Paul C. Silva; Richard L. Moe

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Cryptostomata

An extinct order of Bryozoa in the class Stenolaemata. Cryptostomes had basally attached, erect, calcified colonies made of sheets or delicate branches composed of short tubular or box-shaped feeding zooids (individual units that constitute these colonial organisms) and commonly extrazooidal skeletal deposits. *See* BRYOZOA; STENOLAEMATA.

Morphology. Most cryptostomes have a sharp bend at the transition from the thin-walled inner portion to the thick-walled outer portion of zooids, and this transition point commonly has one or more shelflike structures called hemisepta that generate pronounced constrictions in the width of the zooids. Complete transverse partitions (diaphragms) crossing the interior of zooids may be present. The thick outer walls were built of superposed laminae typically interrupted by small, transverse rod-shaped structures (styles). Overall colony morphology is the basis for dividing cryptostomes into four subgroups, discussed below, that commonly are themselves treated as orders. Small "space-filling" zooids were common, but other types of polymorphic zooids were uncommon.

Classification and history. Rhabdomesina is characterized by arborescent colonies with cylindrical or nearly cylindrical branches, typically having feeding zooids located around the entire perimeter of the branches; some colonies tend to have feeding zooids concentrated toward one side of branches and an extrazooidal skeletal band on the opposite side. Colonies either may be rigidly calcified or may have calcified branch segments joined at flexible, uncalcified joints. Rhabdomesina ranged from the Early Ordovician into the Late Triassic.

Ptilodictyina is characterized by bifoliate colonies in which zooids grew back to back and were oriented toward opposite surfaces of relatively flattened branches or erect sheets. Branches typically divided dichotomously. Closely spaced branching and branch fusion produced reticulated colonies in some ptilodictyines. Others with more widely spaced branching generated arborescent colonies. A few grew as a rigid erect structure above a flexible joint at the base, some others had flexible joints at

branch junctions, but most were completely rigid. Ptilodictyine colonies that grew as sheets had regularly spaced interruptions on the surfaces such as small groups of depressed or protuberant zooids, or spots where feeding zooids were absent. Ptilodictyina existed from the Early Ordovician through the late Carboniferous.

Fenestellina is characterized by colonies with very narrow branches that had feeding zooids oriented toward one surface and massive extrazooidal skeletal deposits on the opposite surface. Branches were closely spaced and were typically linked, either by bars made of extrazooidal skeleton or by anastomosis, into netlike (fenestrate) colonies. Polymorphic zooids were uncommon. These organisms ranged from the Early Ordovician through the late Permian.

Timanodictyina is characterized by cylindrical or bifoliate branches that had very thick walls, containing abundant tiny styles, in outer portions of zooids. They existed from the lower Carboniferous through the late Permian.

Cryptostomes were exclusively marine and lived in a wide variety of relatively low-energy environments on continental shelves and seas. They were important contributors to the growth of organic mounds during the Late Ordovician and early Carboniferous. Some Carboniferous limestones are made predominantly of fenestellid skeletons, and may be petroleum reservoirs where space between the skeletal fragments was not filled by mud or precipitates.

Phylogenetic relationships among the four subgroups are not clear. Probably the Rhabdomesina is paraphyletic and gave rise to the other three groups, each of which apparently is monophyletic.

Frank K. McKinney

Crystal

A solid in which the atoms or molecules are arranged periodically. Within a crystal, many identical parallelepiped unit cells, each containing a group of atoms, are packed together to fill all space (see **illus.**). In scientific nomenclature, the term crystal is usually short for single crystal, a single periodic arrangement of atoms. Most gems are single crystals. However, many materials are polycrystalline, consisting of many small grains, each of which is a single crystal. For example, most metals are polycrystalline. *See* SINGLE CRYSTAL.

In electronics, the term crystal is restricted to mean piezoelectric crystal. Piezoelectric crystals contract or expand under application of electric voltages, and conversely they generate voltages when compressed. They are used for oscillators, pressure sensors, and position actuators. *See* PIEZOELECTRICITY.

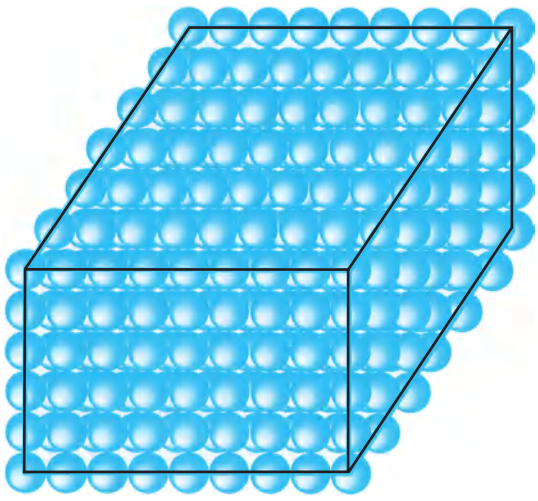
The anisotropic microscopic structure of a crystal is often reflected in its external form, consisting of flat faces and sharp edges. Crystal structure is generally determined via diffraction of x-rays, neutrons, or electrons. Unlike disordered materials such as glasses or liquids, the diffraction pattern of a periodic array of atoms consists of individual sharp spots. The symmetry and structure of the crystal can be inferred from the symmetry of the diffraction pattern and the intensities of the diffracted beams. *See* ELECTRON DIFFRACTION; NEUTRON DIFFRACTION; X-RAY DIFFRACTION.

A crystal can be characterized by the symmetry operations that leave its structure invariant. These can include rotation about an axis through a specific angle, reflection through a plane, inversion through a point, translations by a unit cell dimension, and combinations of these. It can be shown that, for a periodic structure, the only allowable rotational symmetries are twofold, threefold, fourfold, and sixfold (see illus.).

A quasicrystal is a solid which yields a sharp diffraction pattern but has rotational symmetries (such as 5-fold or 10-fold) which are inconsistent with a periodic arrangement of atoms. This apparent paradox can be explained if the structure is assumed to consist of two or more unit cells, arranged according to precise rules. *See* QUASICRYSTAL.

A plastic crystal is generally composed of organic molecules which are rotationally disordered. The centers of the molecules lie at well-defined, periodically spaced positions, but the orientations of the molecules are random. Plastic crystals are often very soft and may flow under their own weight.

A liquid crystal is a material which is intermediate in structure between a liquid and a solid. Liquid crystals usually flow like liquids but have some degree of internal order. They are generally composed of rodlike organic molecules, although in some cases they are composed of disklike molecules. In a nematic liquid crystal, the rods all have the same general orientation, but the positions of the rods are disordered. In a smectic liquid crystal, rodlike molecules are ordered into sheets, within which there is only liquidlike order. A smectic can thus be thought of as being crystalline in one dimension and liquid in



Structure of a simple crystal. Spheres, representing atoms, are packed together into a cubic lattice. This crystal has 4-fold symmetry axes passing through the front face; after a 90° rotation the structure appears unchanged.

the other two. In a discotic liquid crystal, disklike molecules are ordered into columnar arrays; there is short-range liquidlike order within the columns, but the columns form a two-dimensional crystal. *See* CRYSTAL DEFECTS; CRYSTAL GROWTH; CRYSTAL STRUCTURE; CRYSTALLOGRAPHY; LIQUID CRYSTALS.

Paul A. Heiney

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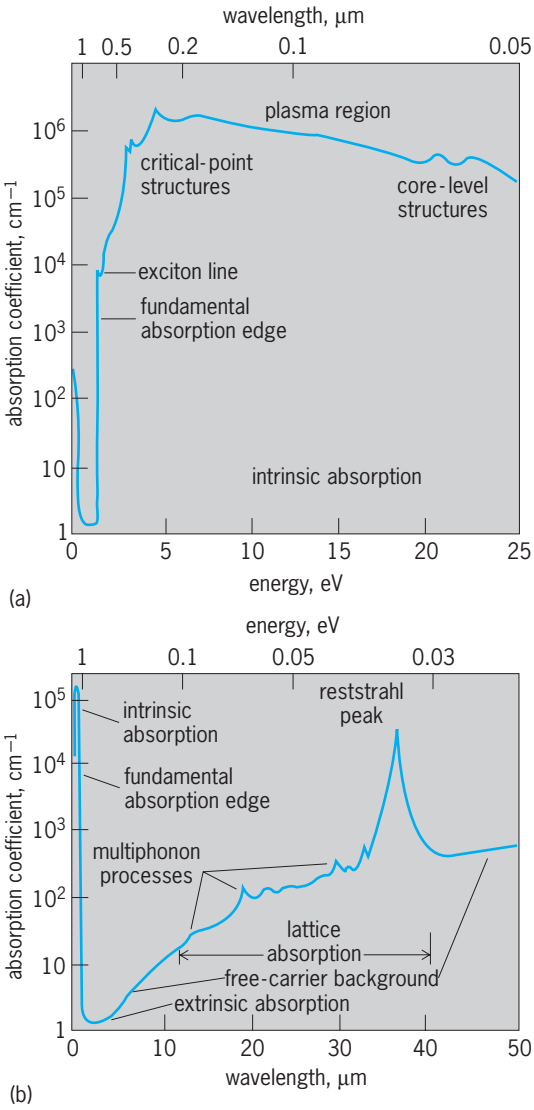
Crystal absorption spectra

The wavelength or energy dependence of the attenuation of electromagnetic radiation as it passes through a crystal, due to its conversion to other forms of energy in the crystal. When atoms are grouped into an ordered array to form a crystal, their interaction with electromagnetic radiation is greatly modified. Free atoms absorb electromagnetic radiation by transitions between a few electronic states of well-defined energies, leading to absorption spectra consisting of sharp lines. In a crystal, these states are broadened into bands, and the cores of the atoms are constrained to vibrate about equilibrium positions. The ability of electromagnetic radiation to transfer energy to bands and ionic vibrations leads to broad absorption spectra that bear little resemblance to those of the free parent atoms. *See* ABSORPTION OF ELECTROMAGNETIC RADIATION; ATOMIC STRUCTURE AND SPECTRA; BAND THEORY OF SOLIDS; LATTICE VIBRATIONS.

The absorption spectrum of a crystal includes a number of distinct features. These are shown in the **illustration** for a semiconducting crystal, gallium arsenide. Absorption spectra of insulating crystals are similar except that electronic absorption is shifted to higher energy. Intrinsic absorption and extrinsic absorption differ according to whether energy would be extracted from electromagnetic radiation even if the crystal were perfect, or if defects or impurities are required. Intrinsic absorption includes lattice absorption in the infrared, electronic absorption by transitions from bonding to antibonding states (valence and conduction bands) in the visible and ultraviolet, and electronic absorption by transitions from core levels in the x-ray region. Extrinsic absorption is much weaker and is generally observable only in the region of transparency of semiconductors and insulators. Free-carrier absorption may be either intrinsic or extrinsic: it is intrinsic to metal and semimetal crystals with their partially filled bands, and extrinsic to semiconductors and insulators with their normally filled or empty bands. Free-carrier processes dominate the absorption spectra of metals from the lowest frequencies throughout the ultraviolet, with lattice and band-to-band absorption processes being much less pronounced. *See* ELECTRIC INSULATOR; METAL; SEMICONDUCTOR.



Absorption spectrum of GaAs. The region in which each type of absorption process is important is indicated. (*a*) Spectrum at higher energies. (*b*) Spectrum at lower energies.

Properties. Absorption is measured in terms of the absorption coefficient, which is defined as the relative rate by which electromagnetic radiation is attenuated per unit length. The absorption coefficient is typically expressed in units of cm^{-1}. Absorption coefficients of solids range from less than 2×10^{-7} cm^{-1} in the exceedingly highly refined glasses used in fiber-optic communications systems to higher than 10^6 cm^{-1} for band-to-band electronic transitions in semiconductors and insulators. Absorption is related to reflection through a more fundamental quantity called the dielectric function. Absorption is also related to the index of refraction by a relationship called the Kramers-Kronig transformation, which can be evaluated if the spectral dependence of the absorption coefficient is known over a sufficiently wide frequency range. *See* DISPERSION RELATIONS; REFLECTION OF ELECTROMAGNETIC RADIATION.

Because energy is extracted from electromagnetic radiation by work (force × distance) performed on the crystal, considerable insight into absorption processes can be obtained from simple mechanical considerations. Although force is just equal to the electric field times the charge, the distance that the charge moves depends inversely on mass and frequency and on the bonding configuration. The mass dependence suggests that absorption associated with the heavy ions of the lattice is much weaker than that associated with the light electrons that hold the lattice together, as can immediately be verified from the illustration. The bonding dependence suggests that absorption associated with the free electrons of a metal such as chromium is different from that associated with the tightly bound electrons of an insulator such as glass, as is immediately verified from observation.

Lattice absorption. Lattice absorption arises from the excitation of phonons, which are collective vibrations of the atomic cores about their equilibrium lattice positions. Lattice absorption occurs in the infrared and is responsible for the intrinsic structure in the absorption spectra of crystals below about 0.1 eV (wavelengths above about 10 micrometers). In a diatomic, partially ionic crystal such as gallium arsenide, an electric field that forces the gallium cores one way simultaneously forces the arsenic cores the other way, leading to the generation of transverse optic phonons. The process becomes very efficient at the natural mechanical resonance of the lattice at $8.8 \times 10^{12}\ s^{-1}$ (energy of 36 meV, wavelength of 34 μm), and causes the strong reststrahl absorption peak in the spectrum of illus. *b*. Since the lattice-restoring forces are anharmonic, it is also possible for electromagnetic radiation to create several phonons simultaneously at higher multiples of the reststrahl frequency. These multiphonon processes are responsible for the structure shown from 10 to 30 μm in illus. *b*. *See* IONIC CRYSTALS; MOLECULAR STRUCTURE AND SPECTRA; PHONON.

Bonding-to-antibonding transitions. Electronic transitions between bonding (filled valence) bands and antibonding (empty conduction) bands yield the crystal analog of the line absorption spectra of free atoms. These intrinsic absorption processes dominate the optical behavior of semiconductors and insulators in the visible and ultraviolet regions of the electromagnetic spectrum, as binding energies of most materials are of the order of magnitude of electronvolts per atom. Although analogous transitions also occur in metals, these are generally overwhelmed by free-carrier effects with the spectacular exceptions of copper, silver, and gold, where free-carrier and band-to-band absorption processes are of comparable importance.

The energy of an electron within a band is not constant but is a function of the velocity or momentum of the electron relative to the crystal lattice. This dependence on momentum causes optical structure within a bonding-to-antibonding absorption band. The most readily apparent such feature is the fundamental absorption edge of semiconductors and insulators, which marks the boundary between the range of transparency at lower energies and the strong electronic absorption that occurs at higher energies. The energy of the fundamental absorption edge is determined by the forbidden gap, which is the energy difference between the highest filled valence and lowest empty conduction level of the crystal. The forbidden gap may be nonexistent, as in metals such as iron, lead, and white tin, and in semimetals such as arsenic, antimony, and bismuth; zero, as in the semiconductors gray tin and one of the mercury-cadmium-telluride alloys; in the infrared, as in the semiconductors onductors germanium, silicon, gallium arsenide, and lead sulfide; in the visible or near-ultraviolet, as in the semiconductors gallium phosphide, cadmium sulfide, and gallium nitride; or in the ultraviolet, as in insulators such as diamond, strontium titanate, sodium chloride, silicon dioxide, and lithium fluoride.

Bonding-to-antibonding absorption processes are of two types, direct or indirect, according to whether the initial and final electronic states have the same momentum or whether the electron must absorb or emit a phonon to conserve momentum in the process. The need to involve phonons makes indirect transitions about 100–1000 times less probable than direct transitions. As absorption coefficients for indirect transitions are correspondingly reduced relative to those for direct transitions, it is not surprising that indirect absorption is typically seen only in semiconducting crystals such as germanium, silicon, gallium phosphide, and aluminum arsenide where the highest valence and lowest conduction band states occur at different momenta. In other semiconducting crystals, for example, indium antimonide, gallium arsenide, and zinc oxide, the highest valence and lowest conduction-band states occur at the same momentum and the fundamental absorption edge is direct.

The detailed dependence of absorption on frequency in the vicinity of the fundamental absorption edge depends not only on the variation of electron energy with momentum but also on many-body effects, that is, interactions between the electron excited into the conduction band and the vacancy, or hole, left behind. Because electrons and holes are charges of opposite sign, there is a strong tendency for an electron and hole to bind together in a state like a hydrogen atom. This state is called an exciton. As a result of the positive binding energy of the exciton, the energy needed to create it is less (by the binding energy) than that of the forbidden gap itself. Consequently, electromagnetic radiation is absorbed at energies less than what would be expected in the absence of electron-hole interactions. The lowest possible exciton absorption process produces the single absorption line seen at the fundamental edges of most semiconductors and insulators at low enough temperatures, and is analogous to the lowest absorption line of a hydrogen atom. Excitonic effects also substantially enhance optical absorption

at energies immediately above the fundamental edge. *See* EXCITON.

Direct thresholds analogous to the fundamental absorption edge also occur at higher energies, and cause additional structure in bonding-to-antibonding absorption spectra. These features are generally described in terms of critical points. In three-dimensional energy bands, four types of critical points occur. Local or absolute minima, including the fundamental absorption edge itself and other thresholds where transitions between new band pairs become possible, are of type M_0. Saddle points where the interband energy reaches local minima in two dimensions and a local maximum in the third are of type M_1. Saddle points where the interband energy reaches a local minimum in one dimension and maxima in the other two are of type M_2. Points where the interband energy reaches a local maximum in all three dimensions are of type M_3.

Critical point structure begins to die out at energies greater than 10 eV, as the crystal potential in which the excited electrons propagate becomes small compared to their kinetic energy. This is the plasma region shown in illus. *a*. Although small groups of absorption features typically appear at higher energies (at 20 eV in illus. *a*), these high-energy structures no longer involve bonding states but are associated with the electronic levels of the atomic cores. Being highly localized, core electrons are virtually unaffected by the crystal and (apart from minor shifts in electrostatic potential) can be viewed as being the same as in the free parent atoms. The core-level features of gallium arsenide seen near 20 eV in illus. *a* originate from the gallium 3*d* core electrons; the 3*d* core electrons of arsenic give rise to similar features (not shown) near 45 eV. Deeper-lying core levels give corresponding absorption features at higher energies. Core-level features in both crystalline and free-atom absorption spectra tend to be broad owing to the short lifetimes of core holes created by the absorption process.

Extrinsic absorption. Extrinsic absorption processes involve states associated with deviations from crystal perfection, such as vacancies, interstitials, and deliberately or inadvertently introduced impurities. Since the concentration of imperfections is typically low relative to the concentration of host atoms (10^{12}–10^{19} cm^{-3} compared to 10^{23} cm^{-3}), extrinsic absorption is weak relative to intrinsic absorption. Consequently, as with indirect absorption, extrinsic absorption is typically seen only in spectral regions where the crystal is otherwise transparent. However, as electromagnetic radiation can interact with many absorbing centers in passing through macroscopic lengths of ordinarily transparent material, extrinsic absorption is an important phenomenon. Extrinsic processes are responsible for the colors of most gemstones, for example, the red of ruby and blue of sapphire, and also for the poor optical quality of industrial diamonds. Extrinsic processes are also vitally important in luminescence, which depends almost entirely on the types and concentrations of impurities in a crystal. *See* LUMINESCENCE.

Free-carrier absorption. Unbound charges such as conduction electrons and valence holes can interact with electromagnetic radiation at all frequencies. However, the absorption of electromagnetic radiation requires that the motion induced by the electromagnetic field be coupled to dissipative mechanisms involving charge-charge or charge-lattice collisions. Thus, although all common metals have large densities of free carriers (of the order of magnitude of 10^{23} cm^{-3}), it is the existence of highly efficient loss mechanisms in transition metals such as iron, nickel, chromium, platinum, and palladium that gives them their characteristic neutral gray appearance. In contrast, the loss mechanisms in aluminum and in the noble metals copper, silver, and gold are much less efficient, leading among other effects to substantially higher reflectances for these materials.

Free-carrier effects are typically described phenomenologically by the simple Drude model with electron density and scattering lifetime as the only two parameters. At frequencies below the reciprocal of the scattering lifetime, absorption is roughly independent of frequency and proportional to the direct-current conductance. Above this value, absorption drops roughly as the square of the frequency. The intrinsic absorption properties of electrons in transition metals and the extrinsic absorption properties of free carriers in semiconductors can both be described approximately by the Drude model. However, the noble metals copper and gold cannot be so described, as a result of the strong interband thresholds in the visible that also give these metals their characteristic colors. Free-electron absorption in high-temperature superconductors may be affected similarly by a strongly frequency-dependent loss mechanism, although the detailed nature of this mechanism is still uncertain. *See* FREE-ELECTRON THEORY OF METALS; SUPERCONDUCTIVITY.

David E. Aspnes

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Crystal counter

A device, more correctly described as a crystal detector, that detects ionizing radiations of all types and is adaptable to measuring neutrons. The sensitive element is a single crystal with a dc resistance normally higher than 10^{12} ohms. The crystals are small and are cut or grown to volumes ranging from less than

1 mm^3 (6×10^{-5} $in.^3$) to approximately 200 mm^3 (1.2×10^{-2} $in.^3$).

Crystal detectors fall into two categories: Certain crystals act as thermoluminescent detectors, of which lithium fluoride (LiF), lithium borate ($Li_2B_4O_7$), and calcium sulfate ($CaSO_4$) are among the best known. Other crystals, for example, cadmium telluride (CdTe) and mercury iodide (HgI_2), act as ionization chambers that deliver either pulses or a dc signal, depending upon the associated electronic circuitry. *See* IONIZATION CHAMBER; THERMOLUMINESCENCE.

Diamond deserves special mention. It is a unique crystal that functions as a thermoluminescent detector or, if suitable contacts are made, as an ionization chamber. The efficiency of the diamond detector in the thermoluminescent or ionization chamber mode is strongly dependent on the impurity atoms included within the crystal lattice, with nitrogen and boron playing dominant roles. Not all diamonds are good detectors; only the rare and expensive natural types IB or IIA are appropriate. Besides being stable and nontoxic, diamond has an additional attractive feature as a detector. As an allotrope of carbon, it has the atomic number $Z = 6$. Human soft tissue has an effective $Z = 7.4$, so that diamond is a close tissue-equivalent material, an essential characteristic for biological dosimetry, for example, in measurements in living organisms. Because of diamond's desirable properties as a detector, economical techniques for the synthesis of diamonds with the correct types and concentrations of impurity atoms are being actively sought for both thermoluminescent and ionization chamber modes. *See* DIAMOND.

Good crystal detectors are insulators and therefore have significant band-gap energies. A large band gap impedes the spontaneous excitation of charge carriers between the valence and conduction bands, thus lowering leakage currents and movement of charge carriers to trapping centers. Room temperature devices are consequently possible. Band gaps for some typical crystal detectors are listed in the **table**. *See* BAND THEORY OF SOLIDS; ELECTRIC INSULATOR; TRAPS IN SOLIDS.

Thermoluminescence. Thermoluminescence can be explained in terms of the band theory of solids. It is to be understood within the framework of the distinctions between fluorescence, which is luminosity continuing only as long as the radiation is continued, and is essentially temperature-independent, and phosphorescence, which is observable luminosity even after the removal of the radiation source.

Band gaps of crystal detectors

Crystal	Band gap, eV	Mode
Diamond	5.4	Thermoluminescence and conduction
LiF	11.0	Thermoluminescence
CdTe	1.45	Conduction
HgI_2	2.15	Conduction

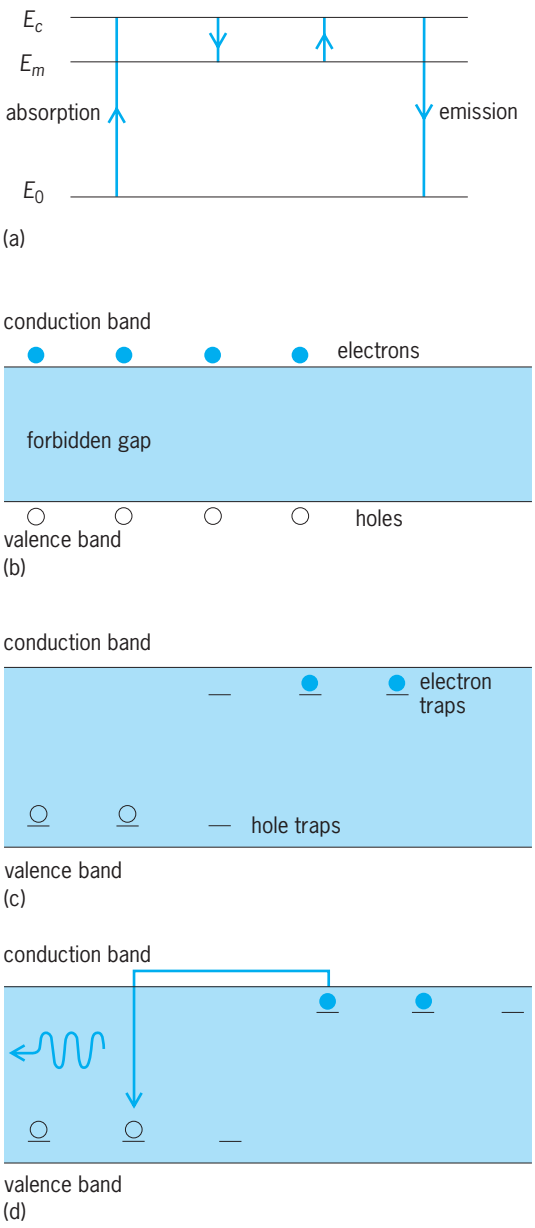

Fig. 1. Crystal band theory of luminescence: (*a*) energy levels; (*b*) conduction and valence bands; (*c*) hole and electron traps; (*d*) photon emission.

Phosphorescence has an intensity decay time which is highly sensitive to the crystal temperature. This situation can arise when the atom or molecule is excited by ionizing radiation—for example, from its ground state E_0 to a metastable state E_m, from which the immediate return to the ground state is prohibited by selection rules (**Fig. 1*a***). If, however, energy $E_c - E_m$ can be supplied to the system, then it can be further excited to a state E_c from which it can return to the ground state directly with the emission of a photon. *See* FLUORESCENCE; LUMINESCENCE; PHOSPHORESCENCE.

Crystals always contain defects in the lattice. Some, such as impurity atoms, can be introduced intentionally or unintentionally during the growing phase. Others are distortions of the lattice. These

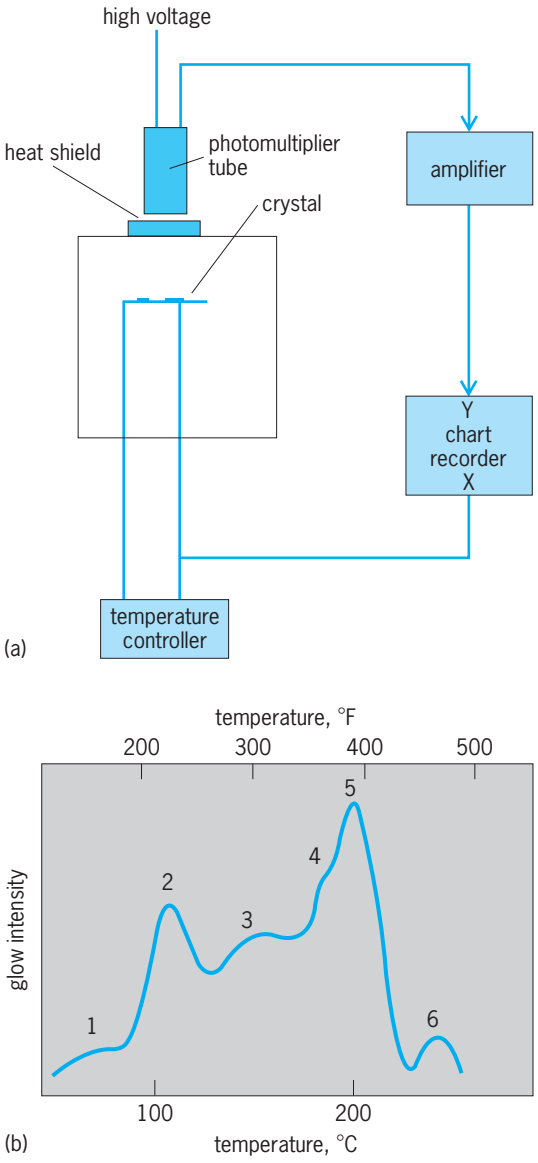

Fig. 2. Thermoluminescence: (*a*) measurement; (*b*) glow curve.

defects perturb the periodicity of the lattice and constitute localized energy levels within the band gap between the valence and conduction bands. After irradiation and excitation of the crystal, the carriers moving through their respective bands (electrons through the conduction band and holes through the valence band; Fig. 1*b*) can become trapped at these defect lattice sites (Fig. 1*c*). Upon heating of the crystal, the trapped electrons (holes) may gain sufficient energy to escape from the trapping centers to the conduction (valence) band, from which they may make a direct transition back to the valence (conduction) band, become retrapped, or combine with a trapped hole (electron). The last alternative, called recombination, can lead to energy release, one form of which is the emission of a photon (Fig. 1*d*). The trapped carriers are known as recombination centers if the emission of photons dominates the energy release of the recombination. The degree of heating required to empty the traps is a function of their depth, and the rate at which they are emptied is a function of the temperature. *See* CRYSTAL DEFECTS; ELECTRON-HOLE RECOMBINATION; HOLE STATES IN SOLIDS.

The crystal is heated at a controlled rate on a metal tray by means of an electric current. The photon emission from the crystal is monitored by a photomultiplier the output of which is directed toward an appropriate recording device (**Fig. 2***a*). The result is a "glow curve" (Fig. 2*b*), the area of which correlates with the number of traps depopulated. These traps were, of course, the traps which were populated during irradiation, their number being directly related to the radiation-field intensity. The integrated light output therefore becomes a direct measure of the total radiation dose. The peaks 1–6 (Fig. 2*b*) are indicative of the various trapping levels within the crystal. *See* DOSIMETER; PHOTOMULTIPLIER.

Conduction mode. An essential feature of the crystal operating as a detector in the conduction mode is the forming of the contacts. The contacts must be ohmic and should not form a Schottky barrier, which would severely impede the charge signal from the crystal. A polarizing voltage is applied, usually to opposite faces of the crystal, and a field of 10^4–10^5 V/cm is used to collect the charges (**Fig. 3**).

A charged particle entering the crystal transfers its kinetic energy to the bulk of the crystal by creating charge carriers (electron-hole pairs). A photon of sufficient energy interacts with the crystal atoms, losing all or part of its energy through the photoelectric effect, the Compton effect, or pair production. In each of these processes, electrons are either liberated or created, and they in turn have their energy dissipated in the bulk of the crystal by creating charge carriers. *See* COMPTON EFFECT; ELECTRON-POSITRON PAIR PRODUCTION; GAMMA-RAY DETECTORS; PHOTOEMISSION.

When the carrier pair is created, the individual carriers move under the influence of the electric field toward the oppositely charged contacts. On arriving at the contacts, the charges can be measured at the output point either as a dc or as a pulse signal, depending upon the circuitry. It is, however, necessary for full

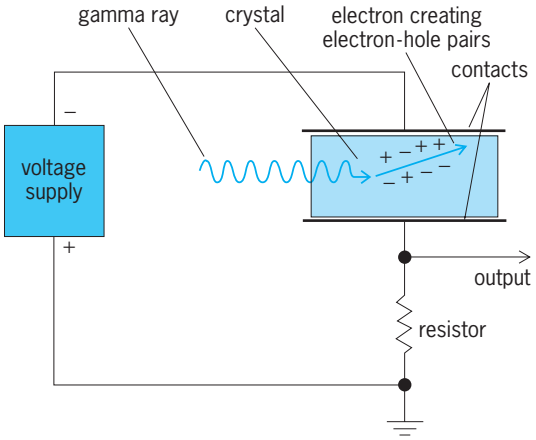

Fig. 3. Crystal detector in conduction mode.

efficiency of the counting system that both types of carriers are collected equally. This makes the property of ion mobility within the crystal very important, and both types should have equal or nearly equal mobility. A condition called hole trapping arises if this criterion is not satisfied. It leads to poor resolution of the overall detecting system, which is also a consequence of high leakage currents or electronic noise within the circuit elements. Because of the high resistance of the crystal, the length of the leads and the screening of all connections must be carefully controlled to reduce the inherent input capacitance of the front end of the system. Resolution can also be improved by cooling the crystal.

Sensitivity and resolution. In the thermoluminescent mode the crystals measure the total dose of the applied radiation, whereas in the conduction mode they measure the instantaneous dose rate; in both cases it is ultimately the crystal itself that limits the sensitivity and resolution of the system. Present methods of synthesis for crystals permit the detection of radiation fields down to nearly background values (0.1 microgray/h or 10^{-5} rad/h) even with crystals as small as 1 mm^3 (6×10^{-5} $in.^3$). Small crystals make detectors possible that are capable of very high spatial resolution. This feature is important in electron radiation therapy, for example, where the isodose curves lie very close together. *See* PARTICLE DETECTOR; RADIATION BIOLOGY; RADIOLOGY.

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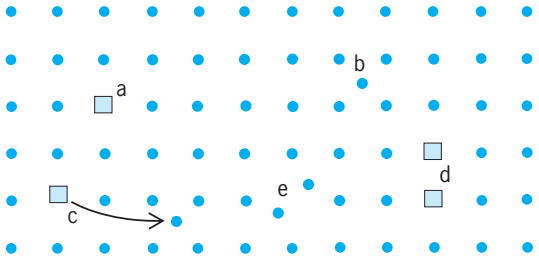
Crystal defects

Departures of a crystalline solid from a regular array of atoms or ions. A "perfect" crystal of sodium chloride (NaCl), for example, would consist of alternating Na^+ and Cl^- ions on an infinite three-dimensional simple cubic lattice, and a simple defect (a vacancy) would be a missing Na^+ or Cl^- ion. There are many other kinds of possible defects, ranging from simple and microscopic, such as the vacancy and other structures (**Fig. 1**), to complex and macroscopic, such as the inclusion of another material, or a surface.

Natural crystals always contain defects, often in abundance, due to the uncontrolled conditions under which they were formed. The presence of defects which affect the color can make these crystals valuable as gems, as in ruby [chromium replacing a small fraction of the aluminum in aluminum oxide (Al_2O_3)]. Crystals prepared in the laboratory will also always contain defects, although considerable control may be exercised over their type, concentration, and distribution.

The importance of defects depends upon the material, type of defect, and properties which are being considered. Some properties, such as density and elastic constants, are proportional to the concentration of defects, and so a small defect concentration will have a very small effect on these. Other properties, such as the color of an insulating crystal or the conductivity of a semiconductor crystal, may be much more sensitive to the presence of small numbers of defects. Indeed, while the term defect carries with it the connotation of undesirable qualities, defects are responsible for many of the important properties of materials, and much of solid-state physics and chemistry and materials science involves the study and engineering of defects so that solids will have desired properties. A defect-free silicon crystal would be of little use in modern electronics; the use of silicon in devices is dependent upon small concentrations of chemical impurities such as phosphorus and arsenic which give it desired electronic properties.



Key:
a = vacancy (Schottky defect)
b = interstitial
c = vacancy-interstitial pair (Frenkel defect)
d = divacancy
e = split interstitial
□ = vacant site

Fig. 1. Some simple defects in a lattice.

Surfaces and interfaces. The importance of surfaces depends very much upon the properties which are being considered and upon the geometry of the specimen. In a specimen no more than a few hundred atoms thick, a significant fraction of the atoms is close to the surface, while in a sample 0.4 in. (10^{-2} m) thick or greater, a much smaller fraction of the atoms is close to the surface. In the latter case, it makes sense to consider bulk properties which are surface-independent and shape-independent. Even in this case, surface effects may be important, as in metals which reflect visible light from their surfaces, but they may generally be separated from bulk effects.

Surfaces are of great importance in determining the chemical properties of solids, since they are in contact with the environment. Such properties include chemical reactions involving the surface, such as corrosion; they also include the role of solids as catalysts in chemical reaction. It is often found that surface atoms reconstruct; that is, they displace and form bonds different from those existing in the bulk. Experimental studies of surfaces are difficult, because of the high vacuum which must be maintained and the relatively small number of atoms

which are involved. *See* CATALYSIS; CORROSION; SURFACE PHYSICS.

An interface is the boundary between two dissimilar solids; it is in a sense an interior surface. For example, a simple metal-oxide-semiconductor (MOS) electronic device has two interfaces, one between metal and semiconductor, and the other between semiconductor and oxide. (In this case the oxide is not crystalline.) Interfaces tend to be chemically and electrically active and to have large internal strains associated with structural mismatch.

Interfaces are of great importance in microelectronics. Modern electronic devices involve very small features and layers of thin semiconducting and insulating films. The success of silicon technology is based in large part on the quality and integrity of the interface between silicon and silicon dioxide. Much of the challenge in developing new generations of semiconductor devices using materials other than silicon lies in finding ways to form high-quality interfaces between different semiconductors and between semiconductors and insulators. *See* INTEGRATED CIRCUITS.

Chemical impurities. An important class of crystal defect is the chemical impurity. The simplest case is the substitutional impurity, for example, a zinc atom in place of a copper atom in metallic copper. Impurities may also be interstitial; that is, they may be located where atoms or ions normally do not exist.

In metals, impurities usually lead to an increase in the electrical resistivity. Electrons which would travel as unscattered waves in the perfect crystal are scattered by impurities. Thus zinc or phosphorus impurities in copper decrease the conductivity. Impurities can play an important role in determining the mechanical properties of solids. As discussed below, an impurity atom can interact with an extended structural defect called a dislocation to increase the strength of a solid; hydrogen, on the other hand, can lead to brittle fracture of some metals. *See* ELECTRICAL CONDUCTIVITY OF METALS; ELECTRICAL RESISTIVITY.

Impurities in semiconductors are responsible for the important electrical properties which lead to their widespread use. The electronic states of all solids fall into energy bands, quasicontinuous in energy. In nonmetals there is a band gap, a region of energy in which no states exist, between filled (valence) and empty (conduction) bands. Because electrons in filled bands do not conduct electricity, perfect nonmetals act as insulators. Impurities (and other defects) often introduce energy levels within the forbidden gap (**Fig. 2**). If these are close in energy to the valence or conduction band, electrons may be thermally excited from a nearby filled level (donor) into the conduction band, or from the valence band into a nearby empty level (acceptor), in the latter case leaving an unoccupied state or hole in the valence band. In both cases, conduction will occur in an applied electric field. Semiconductor devices depend on the deliberate addition (doping) of appropriate impurities. Other impurities may have energy levels which are not close in energy to either

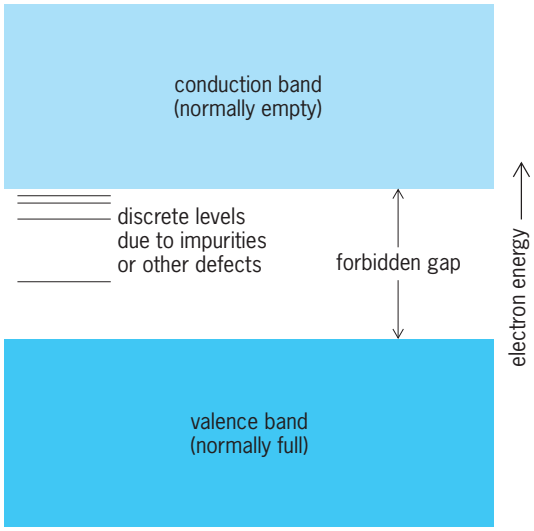


Fig. 2. Band structure of an imperfect nonmetallic crystal, including defect states which occur in the normally forbidden energy gap.

valence or conduction band (so-called deep traps). These may be undesirable if they trap the conducting particles. Thus careful purification is an important part of semiconductor technology. *See* ACCEPTOR ATOM; BAND THEORY OF SOLIDS; DONOR ATOM; HOLE STATES IN SOLIDS; SEMICONDUCTOR; TRAPS IN SOLIDS; ZONE REFINING.

The energy levels associated with impurities and other defects in nonmetals may also lead to optical absorption in interesting regions of the spectrum. Ruby and other gems are examples of this; in addition, the phenomenon of light emission or luminescence is often impurity-related. Solid-state lasers generally involve impurity ions. *See* CRYSTAL ABSORPTION SPECTRA; LASER; LUMINESCENCE.

Structural defects. Even in a chemically pure crystal, structural defects will occur. Thermal equilibrium exists when the Gibbs free energy G is minimized. G is given by Eq. (1), plus a pressure-times-

$$G = E - TS \tag{1}$$

volume term which can often be neglected. Here E is the energy, T is the absolute temperature, and S is the entropy. A perfect crystal has a lower energy E than one containing structural defects, but the presence of defects increases the entropy S. Thus at $T = 0$ K, a perfect crystal is stable, but at higher values of T, the quantity G has its lowest value when structural defects exist. *See* FREE ENERGY.

Crystals may be grown by solidifying the molten substance. If a crystal is cooled slowly (annealed), the number of defects will decrease. However, as the temperature decreases it becomes more difficult for atoms to move in such a way as to "heal" the defects, and in practice there are often more defects than would be expected in thermal equilibrium.

In thermal equilibrium the number of defects at temperatures of interest may be nonnegligible. Application of the methods of statistical mechanics yields

Eq. (2). Here n is the number of defects, N is the

$$\frac{n}{N} = Ae^{-BE/kT} \tag{2}$$

total number of possible atomic sites, E is the energy required to create a defect, and k is Boltzmann's constant. The quantities A and B are dimensionless constants, generally of order 1, whose actual value depends on the type of defect being considered. For example, for vacancies in monatomic crystals such as silicon or copper, $A = B = 1$. E is typically of order 1 eV; since $k = 8.62 \times 10^{-5}$ eV/K, at $T = 1000$ K (1340°F), $n/N = \exp\,[-1/(8.62 \times 10^{-5} \times 1000)] \simeq 10^{-5}$, or 10 parts per million. For many purposes this fraction would be intolerably large, although as mentioned this number may be reduced by slowly cooling the sample. *See* STATISTICAL MECHANICS.

Simple defects. Besides the vacancy, other types of structural defects exist (Fig. 1). The atom which left a normal site to create a vacancy may end up in an interstitial position, a location not normally occupied. Or it may form a bond with a normal atom in such a way that neither atom is on the normal site but the two are symmetrically displaced from it. This is called a split interstitial. The name Frenkel defect is given to a vacancy-interstitial pair, whereas an isolated vacancy is a Schottky defect. (In the latter case the missing atom may be thought of as sitting at a normal site at the surface.)

Simple defects are often combined. For example, when a small amount of alkaline-earth halide is melted with an alkali halide and the resulting mixture is recrystallized, the mixed solid is found to have a large number of alkali-ion vacancies, approximately one for each alkaline-earth ion. Since the alkaline-earth ions are divalent, the alkali-ion vacancies act as "charge compensators" to make the crystal neutral. Such a crystal has a relatively high electrical conductivity associated with the movement of positive ions into the vacancies (or, alternatively, movement of the vacancies through the crystal). This general phenomenon of ionic conductivity is of practical interest, for example, in attempts to develop solid-state batteries. *See* IONIC CRYSTALS; SOLID-STATE BATTERY.

Simple defects tend to aggregate if thermally or optically stimulated. The *F*-center in an alkali halide is an electron trapped at a halogen-ion vacancy. Under suitable conditions of optical or thermal stimulation, it is found that *F*-centers come together to form F_2- or *M*-centers (two nearest *F*-centers), F_3- or *R*-centers (three nearest *F*-centers), and so on. *See* COLOR CENTERS.

Extended defects and mechanical properties. The simplest extended structural defect is the dislocation. An edge dislocation is a line defect which may be thought of as the result of adding or subtracting a half-plane of atoms (**Fig. 3***a*). A screw dislocation is a line defect which can be thought of as the result of cutting partway through the crystal and displacing it parallel to the edge of the cut (Fig. 3*b*). The lattice displacement associated with a dislocation is called the Burgers vector $\vec{b}$.

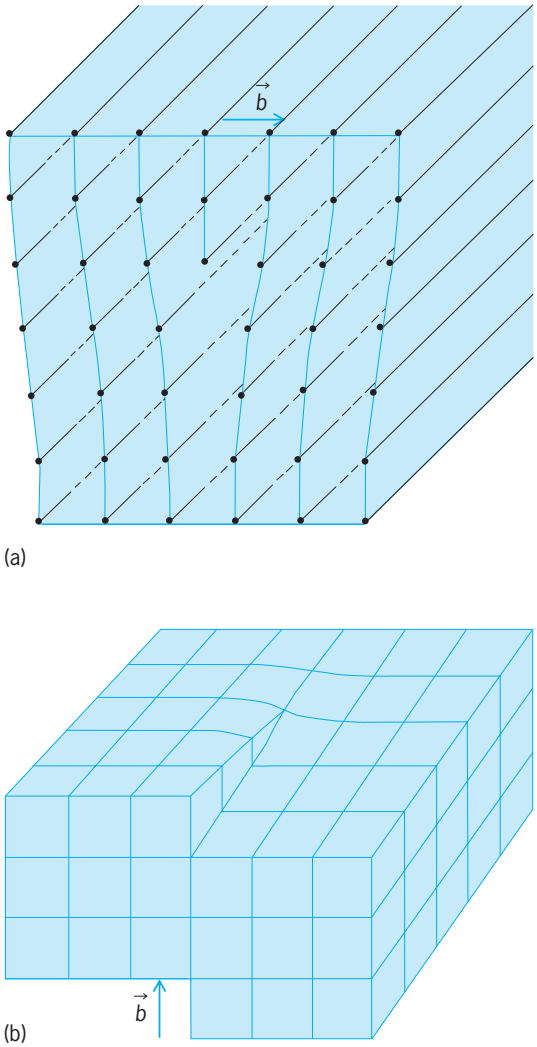


Fig. 3. Extended structural defects. (*a*) Edge dislocation. (*b*) Screw dislocation. In each case the Burgers vector is denoted by $\vec{b}$.

Dislocations generally have some edge and some screw character. Since dislocations cannot terminate inside a crystal, they either intersect the surface or an internal interface (called a grain boundary) or form a closed loop within the crystal.

Dislocations are of great importance in determining the mechanical properties of crystals. A dislocation-free crystal is resistant to shear, because atoms must be displaced over high-potential-energy barriers from one equilibrium position to another. It takes relatively little energy to move a dislocation (and thereby shear the crystal), because the atoms at the dislocation are barely in stable equilibrium. Such plastic deformation is known as slip. *See* PLASTIC DEFORMATION OF METAL.

To extend the earlier argument about the difficulty in preparing perfect crystals, it should be anticipated that most crystals will contain dislocations in ample numbers and that special care would have to be taken to prepare a dislocation-free crystal. The latter is in most cases impractical (although in fact some electronic materials are dislocation-free). Although practical metallurgy developed centuries before

dislocations were known to exist, most strengthening processes are methods for dealing with dislocations.

Dislocations will not be mechanically deleterious if they cannot move. Three major ways of hindering dislocation motion are pinning by dissolved impurities, blocking by inclusions, and tangling. Each of these is described briefly below.

Impurities tend to collect near dislocations if the solid is sufficiently hot that the impurities can move. The presence of these impurities along a dislocation removes the energetic equivalence of the dislocation in different positions; now it requires energy to move the dislocations unless the cloud of dissolved impurities can move as well; and to move them will involve overcoming a large energy barrier. This approach's success is lessened at higher temperatures, when the impurities tend to migrate away.

Small particles of a second substance can "block" dislocation motion. Steel is iron with small inclusions of iron carbide, whose major function is dislocation blocking.

A large, randomly oriented collection of dislocations can tangle and interfere with their respective motions. This is the principle behind work-hardening.

Dislocations have other important properties. The growth of crystals from the vapor phase tends to occur where screw dislocations intersect the surface, for in these regions there are edges rather than planes, and the attractive forces for atoms will be greater. Thus, spiral growth patterns are derived from screw dislocations (**Fig. 4**). *See* CRYSTAL GROWTH.

Acommon planar structural defect is the stacking fault. This can be thought of as the result of slicing a crystal in half, twisting the two pieces with respect to each other, and then reattaching them. Another planar defect, the grain boundary, is the boundary between two crystallites. In some cases it can be described in terms of a parallel set of edge dislocations. *See* TWINNING (CRYSTALLOGRAPHY).

Alloys. A solid containing more than one element is called an alloy, especially when the solid is metallic. There are a vast number of possible alloys, and the present discussion will treat two-component or binary systems as examples.

The term stoichiometry is used to describe the chemical composition of compounds and the resulting crystal structure. The alkali halides are particularly simple, forming compounds with well-defined compositions (for example, NaCl). The situation becomes more complicated for compounds in which one or more components can take on more than one valence state. In particular, in the transition-metal oxides different compositions can occur, with different crystal structures and defects. For example, titanium dioxide (TiO_2) represents a well-defined perfect crystal, whereas TiO_{2-x} would have an oxygen deficiency whose value is described by the subscript x. In this case the deficit oxygens are associated with neither oxygen vacancies nor titanium interstitials, but rather with defect aggregates called shear planes.

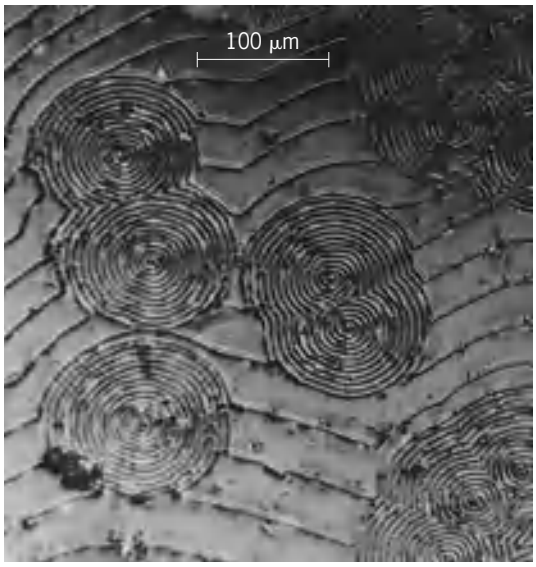


Fig. 4. Crystal-growth spirals on the surface of a silicon carbide crystal. Each spiral step originates in a screw dislocation defect. From the side the surface looks like an ascending ramp wound around a flat cone. (*General Electric Co.*)

If two elements are mixed uniformly throughout the solid, the atomic arrangement may be described as ordered or disordered. For example, an alloy composed of equal numbers of two metal atoms may form an ordered lattice with a regular periodic arrangement of the two atoms with respect to each other, or a disordered lattice in which some of the atoms are on the "wrong" site (**Fig. 5**). The ordered arrangement is most stable at very low temperatures. The temperature at which the structure becomes disordered is called the transition temperature of the order-disorder transformation. *See* ALLOY STRUCTURES.

Defect chemistry. Most of the preceding discussion has treated the simpler types of defects and their static properties, that is, their properties under conditions such that the nature and location of the defects do not change with time. Situations have also been alluded to in which the number, locations, or types of defects change with time, either because of thermal effects (for example, slowly cooling a hot sample and annealing vacancies) or because of

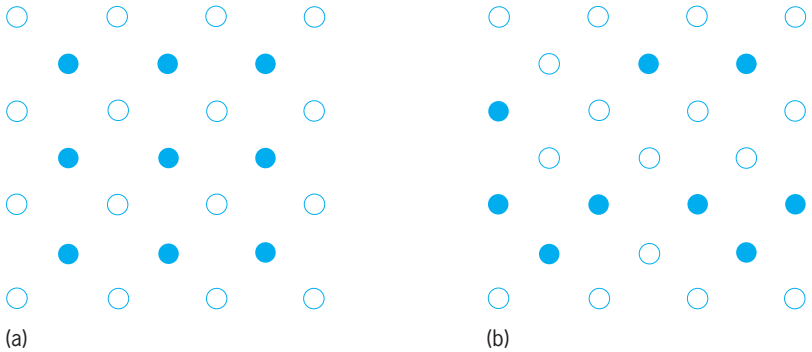


Fig. 5. Binary alloy. (*a*) Ordered lattice. (*b*) Disordered lattice.

external irradiation or defect interactions (causing, for example, the aggregation of alkali-halide *F*-centers).

For both scientific and practical reasons, much of the research on crystal defects is directed toward the dynamic properties of defects under particular conditions, or defect chemistry. Much of the motivation for this arises from the often undesirable effects of external influences on material properties, and a desire to minimize these effects. Examples of defect chemistry abound, including one as familiar as the photographic process, in which incident photons cause defect modifications in silver halides or other materials. Properties of materials in nuclear reactors is another important case. A few examples will be considered here.

Thermal effects. Thermal effects are the easiest to understand from a microscopic point of view: raising the temperature increases the amplitude of atomic vibrations, thereby making it easier for atoms to "hop" from place to place over potential-energy barriers. Almost all dynamic processes are temperature-dependent, occurring more readily at high temperatures. *See* DIFFUSION; LATTICE VIBRATIONS.

Defect creation by irradiation. Radiation can have profound effects on materials in both direct and indirect ways. For example, a beam of high-energy electrons, neutrons, or atomic nuclei (such as alpha particles) can create defects by simple collision processes: the incident particle imparts a portion of its momentum and energy to a lattice atom, thereby releasing the atom from its normal postion. If the released atom has enough energy, it may collide with and release a second lattice atom, thereby creating a second vacancy, and so on; thus one incident particle may lead to the production of a number of defects. Other processes also exist. *See* CHARGED PARTICLE BEAMS; RADIATION DAMAGE TO MATERIALS.

A beam of photons may also create defects through a variety of processes. If the photon energy equals or exceeds the band gap, ionization can occur; that is, electrons can be excited from core or valence bands into the conduction bands as the photons are absorbed. Very energetic photons (gamma rays) may be absorbed or scattered, generating fast electrons, which in turn eject atoms through collision processes such as those occurring with incident particles. *See* GAMMA RAYS.

Lower-energy photons (x-rays, ultraviolet, visible, and so on) have insufficient energy to create defects by such processes. In some solids simple ionization results in defect creation; for example, in alkali halides excitation of an electron on a halogen ion leads after atomic rearrangements to the formation of an excited halogen molecule-ion which is unstable with respect to a translation along its axis. This leads to the creation of a halogen ion vacancy plus an electron (*F*-center) next to a molecule-ion which is crowded into a normal halogen ion position (*H*-center). A replacement sequence can occur (**Fig. 6**) in which each halogen displaces its neighbor in the indicated direction. Eventually this replacement sequence ends, and the result is a pair of isolated *F*- and

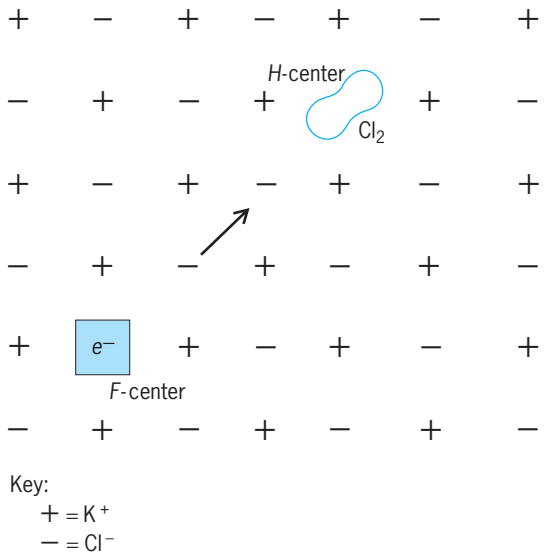


Fig. 6. *F*- and *H*-center formation in KCl after a replacement sequence of halogen-halogen collisions along the direction indicated by the arrow.

H-centers. The first stages of this process can occur in very short times (on the order of 10^{-12} s).

Laser irradiation is often observed to result in crystal destruction, even when the crystal is an insulator and the photon energy is well below the band gap. One way this can occur is first by ionization of defects by the laser, producing a few conduction electrons. These conduction electrons can in turn be accelerated, by the large electric field present in the laser beam, to sufficiently high energy to ionize normal atoms; thus more electrons are available to be accelerated, and a type of cascade process occurs which leads to destruction of the crystal. *See* LASER-SOLID INTERACTIONS.

Effects of radiation on preexisting defects. More subtle effects must often be considered when studying the effects of radiation on device performance. This involves the interaction of radiation with existing (and desirable) defects. Aggregation of *F*-centers occurs when alkali halides are irradiated in the *F*-absorption band. The process involves several steps: (1) The excited *F*-center loses its electron to the conduction band, leaving behind a positively charged vacancy. (2) The conduction electron is subsequently trapped at another *F*-center, forming a (negative) *F*′-center. (3) Under the attractive Coulomb force, the vacancy moves toward the *F*′-center. (4) The extra electron of the *F*′-center returns to the vacancy either by ionization or by tunneling, leading to two *F*-centers closer than normal. (5) The process repeats, until the two *F*-centers are adjacent.

Externally activated defect processes are also found in semiconductors, typically (but not solely) associated with irradiation. Charge-state processes involve the motion of preexisting defects following a change in the charge of the defect. For example, band-gap light generates electrons and holes in silicon, which can be trapped, for example, at a vacancy *V* in silicon. The doubly negative vacancy (V^{2-})

has a considerably smaller activation energy for motion than less negatively charged vacancies, and will consequently migrate more readily. Presumably differences in locations of the neighboring atoms are responsible for the different activation energies of motion.

Recombination processes may also occur. Here an electron and hole recombine at a defect and transfer their (electronic) energy to vibrational energy of the defect, leading to its motion. This effect has been observed, for example, at a recombination junction in gallium arsenide (GaAs), where defects introduced by proton bombardment were destroyed under charge injection. *See* ELECTRON-HOLE RECOMBINATION.

A third possibility is excited-state effects. In this case a defect in an excited electronic state and its surroundings relax into a configuration in which defect motion is easier; in some instances, there may be a zero barrier energy to overcome. One of the best examples of this is the type II F_A-center in alkali halide crystals, an *F*-center with an alkali impurity (usually Li) as one of its six nearest alkali neighbors (**Fig. 7**). After a photon is absorbed, the system relaxes into a "double-well" configuration with a Cl^- which was initially adjacent to both the Li^+ and the vacancy moving into a "saddle-point" position half the way to the vacancy. The excited electron is then shared by the two small equivalent "wells." When the system returns to the ground state, there is equal probability of the Cl^- returning to its original site or to the original vacancy; in the latter case the vacancy will have moved. Similar processes are thought to occur in semiconductors.

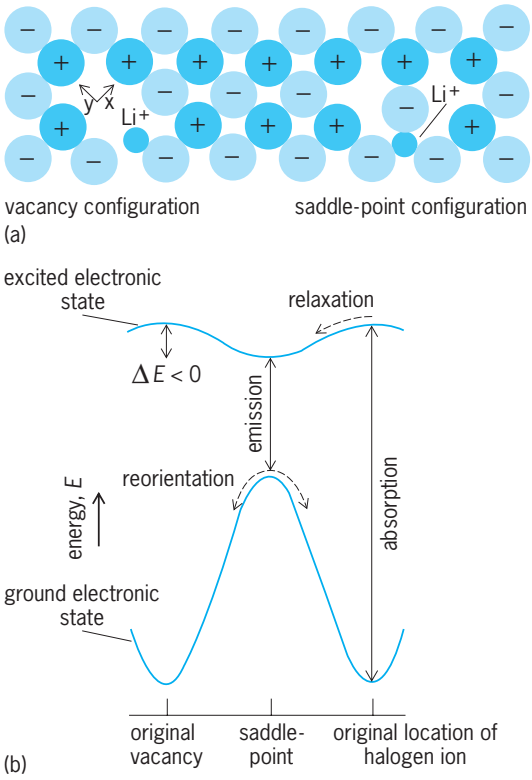


Fig. 7. Reorientation of the type II F_A-center in alkali halides. (*a*) Vacancy and saddle-point configurations. (*b*) Schematic potential energy functions for motion of the neighboring (halogen)$^-$ ion. (*After W. Beall Fowler, ed., Physics of Color Centers, Academic Press, 1968*)

Bistability and metastability. Many semiconductor defects are found to exhibit bistability or metastability. An example is the defect formed in silicon from an interstitial carbon–substitutional carbon pair. This defect has three stable charge states: it can be neutral, it can trap an electron (−), or it can give up an electron (+). Furthermore, it is found to have two distinctly different atomic configurations (whose structures have been proposed but are not known with certainty). Configuration A is stable when the defect is in either the (+) or (−) charge state, while configuration B is stable when the defect is neutral. This phenomenon is called bistability. In addition, in each case the other configuration also has a potential-energy minimum but at a slightly higher energy; this is an example of metastability. Thus when A is stable, B is metastable, and vice versa. At low temperature a metastable state can have a long lifetime if the energy barrier to atomic rearrangement is sufficiently high.

Negative U. Some semiconductor defects possess a characteristic called negative U. This term comes about because the defect behaves as if the electron–electron interaction (U) were attractive. An example is interstitial boron in silicon: this defect can exist in either (−) or (+) charge states, but the neutral state is not stable. This means that more energy is gained when a neutral boron traps an electron than was lost when that electron was removed from another neutral boron: the total energy of the system is lower if half the defects are (+) and half (−) than if all are neutral. In chemistry this phenomenon is known as disproportionation. The occurrence of negative U in semiconductors involves several factors: First, the net Coulomb repulsion between defect electrons is reduced by the interactions with other electrons in the solid (polarization and hybridization). Second, the energies gained from atomic rearrangements can sometimes exceed the remaining Coulomb repulsions, thereby favoring the more ionic charge states.

W. Beall Fowler

Topological defects. Defects such as vacancies and interstitials have been known for many years, as have topological defects such as dislocations. The study of two-dimensional systems and their associated defects has led to the discovery of a new phase of matter which has no crystalline positional order and behaves like a fluid under shear, but has well-defined axes like a crystal. An example is the layered liquid-crystal compound 650 BC, which has properties intermediate between a liquid and a crystal.

Order in crystals. To understand topological defects, it is first necessary to understand the nature of order in crystals that have both translational and bond-orientational order. Translational order means that the atoms of the crystal form a well-defined periodic lattice, while orientational order is weaker and means that the directions of the bonds between neighboring atoms are the same throughout the

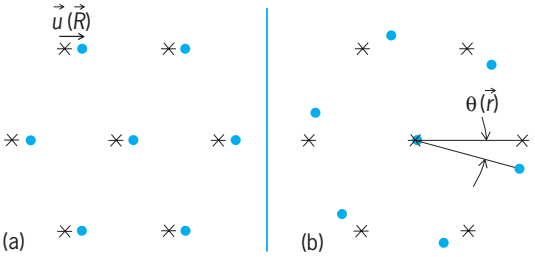


Fig. 8. Translations and rotations of a crystal. The ideal positions are denoted by crosses. (*a*) A patch of crystal translated by $\vec{u}$ from its ideal position with no change of the crystal axes. (*b*) A patch rotated by an angle θ relative to the perfect crystal.

system. A translation of a patch of crystal from its ideal position will spoil translational but not orientational order (**Fig. 8***a*), whereas a rotation will disrupt both (Fig. 8*b*). In principle, orientational order may exist in the absence of translational order, but not vice versa, and a system of point atoms or spherical molecules can exist in three phases: a fully ordered crystal, an intermediate anisotropic fluid with orientational order only, and an isotropic fluid with neither. Systems of nonspherical molecules such as liquid crystals may have many more complex phases. In the great majority of real three-dimensional systems, only the first and last possibilities are realized, with an abrupt melting or freezing transition between states of complete order and complete disorder. However, the intermediate phase with long-range orientational but short-range translational order corresponds to the hexatic phase observed in various liquid-crystal systems and also in xenon adsorbed on graphite. The name hexatic comes from the underlying hexagonal structure of the parent crystal. The three phases can be distinguished by diffraction measurements.

Dislocations and disclinations. A topological defect is one that disrupts a particular order. A crystal, which has two types of order, has two types of topological defect. A translational defect is a dislocation (Fig. 3), and an orientational defect is a disclination. A dislocation may be regarded as a disclination dipole (**Fig. 9**). Circling around the $+\pi/2$ ($-\pi/2$) disclination in a square lattice requires making five (three) right-angle turns and the orientation of the axes is quickly lost. Circling around the dipole involves four right angles, and so this does not disrupt orientational order. However, an attempt to go around the dipole while keeping count of the number of lattice spacings on a path which would close if the lattice were perfect results in failure to return to the starting point by the Burgers vector $\vec{b}$. In other words, the dislocation disrupts translational order only, while a disclination disrupts both translational and orientational order.

Two-dimensional systems. Topological defects do not occur to any significant extent in most bulk crystals in thermal equilibrium. They are present in general as frozen-in metastable states and have a dramatic effect on mechanical properties. In films a few atoms thick and in three-dimensional systems of very weakly coupled layers, they are, however, present in thermal equilibrium and lead to some unusual effects. A graphic way of visualizing a two-dimensional crystal and its excitations and defects is to imagine the crystal lattice drawn on an elastic sheet. This can be distorted in a number of ways, all of which are present to some extent depending on the energy involved, as in Eq. (2). There are the obvious smooth distortions created by stretching and compressing different regions without cutting. Interstitials and vacancies result from the addition or removal of small pieces containing a single atom and regluing the edges. Dislocations are introduced by cutting the sheet, displacing the edges by one lattice spacing, and rejoining the edges, after adding or removing some material as necessary. In a real system this requires diffusing in interstitials or vacancies. Disclinations involve the addition or removal of an angular segment contained between two crystal axes and clearly require a huge amount of energy in the crystalline state.

The equilibrium crystalline state is one with some smooth distortions, some vacancies and interstitials, and a small density of dislocation pairs of equal and opposite Burgers vectors (**Fig. 10***a*). The essential ideas of the role played by topological defects can be understood by focusing on the members of a single

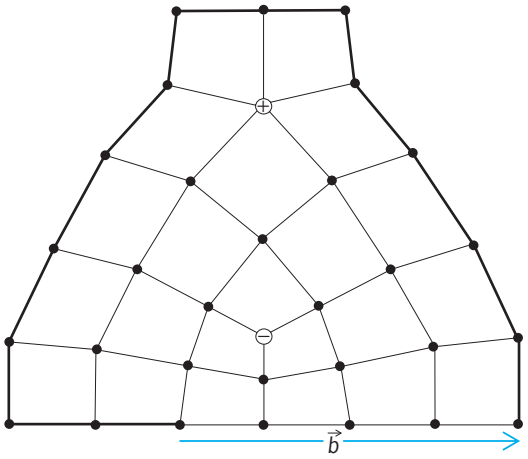


Fig. 9. Disclination dipole in a square lattice. The atom denoted by $\oplus$ ($\ominus$) has five (three) neighbors and corresponds to a $+\pi/2$ ($-\pi/2$) disclination. The heavy line is a perfect lattice closed path.

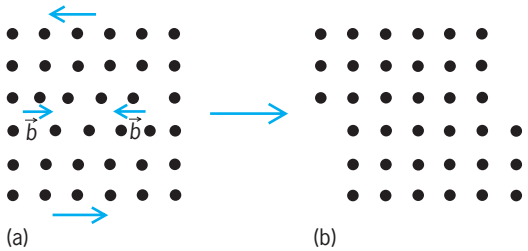


Fig. 10. Defect unbinding transition in a two-dimensional crystal. (*a*) A dislocation pair of equal and opposite Burgers vectors $\vec{b}$ in a square lattice. Above the melting temperature, an infinitesimal shear stress, indicated by the arrows above and below the lattice, will cause the dislocations to move and the crystal to distort. (*b*) Result of this motion.

bound pair. As the temperature is raised, the average spacing between them increases until it equals the spacing between pairs. The density of dislocations remains small. At this stage, the dislocations become independent of each other, positional order is lost, and no force is required to move them about the crystal. The system will behave exactly like a very viscous fluid under a shear stress (Fig. 10*b*).

Provided the density of dislocations and point defects is sufficiently low, the resulting fluid has orientational order. As the temperature is further increased, pairs of equal and opposite disclinations appear in the fluid and their separation will eventually diverge. Above this temperature, all orientational order is lost and the fluid is isotropic.

This discussion assumes that the density of defects remains quite low; in that case the melting of the crystal is via two stages of defect unbinding transitions of the Kosterlitz-Thouless type. Should the density of local defects, such as vacancies, interstitials, and closely bound dislocation pairs, be large, then the crystal melts by an explosive proliferation of these directly to the isotropic fluid. This is the route taken by the great majority of melting transitions both in the bulk and in films, but the defect unbinding mode has been observed experimentally. The theoretical discovery of the hexatic phase led to the experimental identification of the bulk hexatic B and smectic I and F liquid-crystal phases. *See* CRYSTAL STRUCTURE; CRYSTALLOGRAPHY; LIQUID CRYSTALS.

John M. Kosterlitz

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Crystal growth

The growth of crystals, of which all crystalline solids are composed. This article discusses the processes by which crystal growth occurs, and methods of artificially growing crystals.

Growth Processes

Crystal growth generally comes about by means of the following processes occurring in series: (1) diffusion of the atoms (or molecules, in the case of molecular crystals such as hydrocarbons or polymers) of the crystallizing substance through the surrounding environment (or solution) to the surface of the crystal, (2) diffusion of these atoms over the surface of the crystal to special sites on the surface, (3) incorporation of atoms into the crystal at these sites, and (4) diffusion of the heat of crystallization away from the crystal surface. The rate of crystal growth may be limited by any of these four steps. The initial formation of the centers from which crystal growth proceeds is known as nucleation. *See* CRYSTALLIZATION; NUCLEATION.

Increasing the supersaturation of the crystallizing component or increasing the temperature independently increases the rate of crystal growth. However, in many physical situations the supersaturation is increased by decreasing the temperature. In these circumstances the rate of crystal growth increases with decreasing temperature at first, goes through a maximum, and then decreases. Often the growth is greatly retarded by traces of certain impurities.

If nucleation occurs in several locations in the medium, the crystals grow isolated from one another for a time. However, if several differently oriented crystals are growing, they may finally impinge on one another, and intercrystalline (grain) boundaries will be formed. At relatively high temperatures, the average grain size in these polycrystalline aggregates increases with time by a process called grain growth, whereby the larger grains grow at the expense of the smaller.

Regular growth. During its growth into a fluid phase, a crystal often develops and maintains a definite polyhedral form which may reflect the characteristic symmetry of the microscopic pattern of atomic arrangement in the crystal. The bounding faces of this form are those which are perpendicular to the directions of slowest growth. How this comes about is illustrated in **Fig. 1**, in which it is seen that the faces *b*, normal to the faster-growing direction, disappear, and the faces *a*, normal to the slower-growing directions, become predominant. Growth forms, like that shown in the figure, are not necessarily equilibrium forms, but they are likely to be most regular when the departures from equilibrium are not large. *See* CRYSTAL STRUCTURE.

The atomic binding sites on the surface of a crystal can be of several kinds. Thus an atom must be more weakly bound on a perfectly developed plane of atoms at the crystal surface (site A in **Fig. 2**) than at a ledge formed by an incomplete plane one atom thick (site B). Atom A binds with only three neighboring atoms, whereas atom B binds to five neighbors.

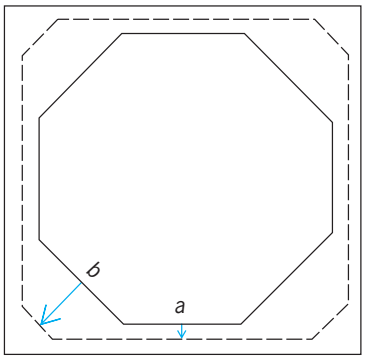

Fig. 1. Schematic representation of cross section of crystal at three stages of regular growth.

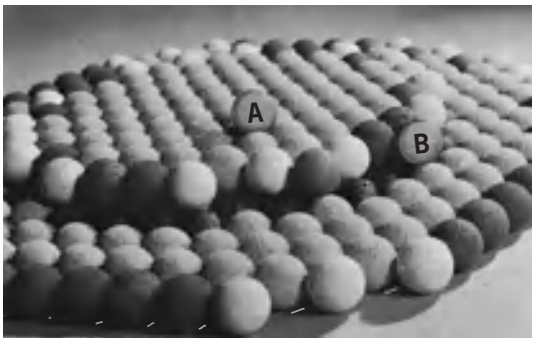

Fig. 2. A model of a crystal surface showing the inequivalent binding sites at which the atoms labeled A and B are located. (*General Electric Co.*)

(An atom in the middle of the island monolayer has bonds with nine neighbors.) Therefore, the binding of atoms in an island monolayer on the crystal surface will be less per atom than it would be within a completed surface layer.

The potential energy of a crystal is most likely to be minimum in forms containing the fewest possible ledge sites. This means that, in a regime of regular crystal growth, dilute fluid, and moderate departure from equilibrium, the crystal faces of the growth form are likely to be densely packed and atomically smooth. There will be a critical size of monolayer, which will be a decreasing function of supersaturation, such that all monolayers smaller than the critical size tend to shrink out of existence, and those which are larger will grow to a complete layer. The critical monolayers form by a fluctuation process. Kinetic analyses indicate that the probability of critical fluctuations is so small that in finite systems perfect crystals will not grow, except at substantial departures from equilibrium. That, in ordinary experience, finite crystals do grow in a regular regime only at infinitesimal departures from equilibrium is explained by the screw dislocation theory. According to this theory, growth is sustained by indestructible surface ledges which result from the emergence of screw dislocations in the crystal face. *See* CRYSTAL DEFECTS.

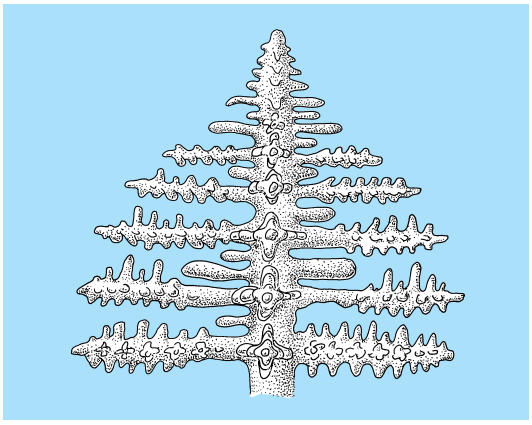

Fig. 3. A dendrite. Growth direction is upward for primary protuberances; side branches grow perpendicularly; side branches on the side branches are also seen growing perpendicularly. (*After W. Kurz and D. J. Fisher, Fundamentals of Solidification, Trans Tech Publications, 1986*)

Irregular growth. When the departures from equilibrium (supersaturation or undercooling) are sufficiently large, the more regular growth shapes become unstable and cellular (grasslike) or dendritic (treelike) morphologies develop. Essentially, the development of protuberances on an initially regular crystal permits more efficient removal of latent heat or of impurities, but at the cost of higher interfacial area and the associated excess surface energy. In the cellular structure, the protuberances align, forming an advancing front and rejecting impurities laterally into the spaces between protuberances. In the dendritic morphology, these protuberances become narrower and develop side branches, which are other protuberances growing laterally (**Fig. 3**). When the supersaturation becomes so great that the energy associated with the increase in interfacial area is unimportant, protuberances proliferate and the crystal grows in a multibranched form that is even more complicated; its shape is characterized by fractal geometry. This limit, in which atoms stick permanently wherever they land, is called diffusion-limited aggregation because steps 2–4 in the series of processes outlined above occur very rapidly, and the rate of crystal growth is limited by diffusion (step 1). *See* FRACTALS.

Michael J. Aziz; David Turnbull

Growth Methods

The advent of semiconductor-based technology generated a demand for large, high-quality single crystals, not only of semiconductors but also of associated electronic materials. With increasing sophistication of semiconductor devices, an added degree of freedom in materials properties was obtained by varying the composition of major components of the semiconductor crystal over very short distances. Thin, multilayered single-crystal structures, and even structures that vary in composition both normal and lateral to the growth direction, are often required.

Bulk single-crystal growth. Bulk single crystals are usually grown from a liquid phase. The liquid may have approximately the same composition as the solid; it may be a solution consisting primarily of one component of the crystal; or it may be a solution whose solvent constitutes at most a minor fraction of the crystal's composition. **Figure 4** shows the phase diagram and corresponding component partial pressures for a hypothetical binary compound. The stoichiometry range of Fig. 4*a* results from the deviation from the 1-to-1 composition caused by point defects in the crystal structure. Nonstoichiometry, high partial pressures, and relatively high point-defect concentrations compared to growth at lower temperatures are all disadvantages of melt growth. Nevertheless, melt growth is used for virtually all bulk semiconductor crystals and for many other industrial crystals. *See* NONSTOICHIOMETRIC COMPOUNDS; PHASE EQUILIBRIUM.

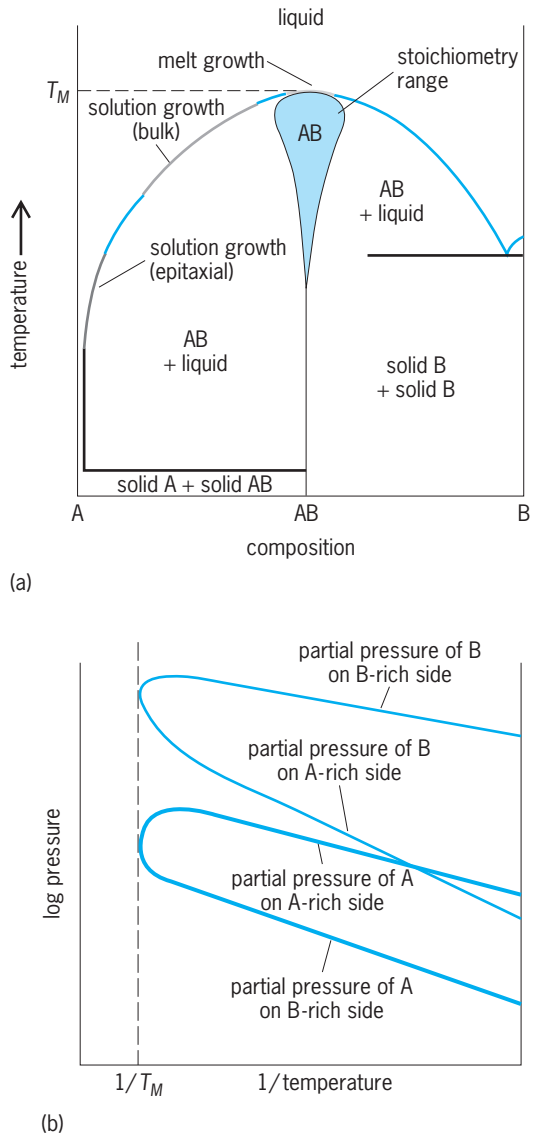

Fig. 4. Hypothetical binary system AB having simple compound formation. (*a*) Phase diagram. (*b*) Corresponding partial pressure curves. The extent of the stoichiometry range is greatly exaggerated for illustration.

Czochralski method. The most important bulk crystal growth technique is the crystal-pulling or Czochralski method, in which a rotating seed crystal is dipped into the melt (**Fig. 5**). Rotation reduces radial temperature gradients, and slow withdrawal of the rotating seed results in growth of a cylinder of single-crystal material. The conditions for optimum growth vary widely, and pulling rates range up to a few inches per hour. Crystal diameter and length depend upon the details of the temperature and pulling rate, and the dimensions of the melt container. Crystal quality depends very critically upon minimization of temperature gradients that enhance the formation of dislocations. Pulled silicon crystals 6 in. (15 cm) in diameter are important for the semiconductor industry. Ruby, sapphire, and group 13–15 compound semiconductor crystals are among the many crystals that are routinely grown by the Czochralski technique.

Melt-growth variations. There are several other variations of the melt-growth technique. One of them is the zone-refining method, in which a narrow molten region is passed through a solid rod of single-crystal or polycrystalline material. Another is the Bridgman technique, in which single-crystal growth is achieved for some materials by moving the melt through a temperature gradient (**Fig. 6**). Generally the container has a restricted region where the first solidification takes place as the liquidus temperature moves past it. An initially formed crystallite in that region propagates with the moving liquidus front, and a single

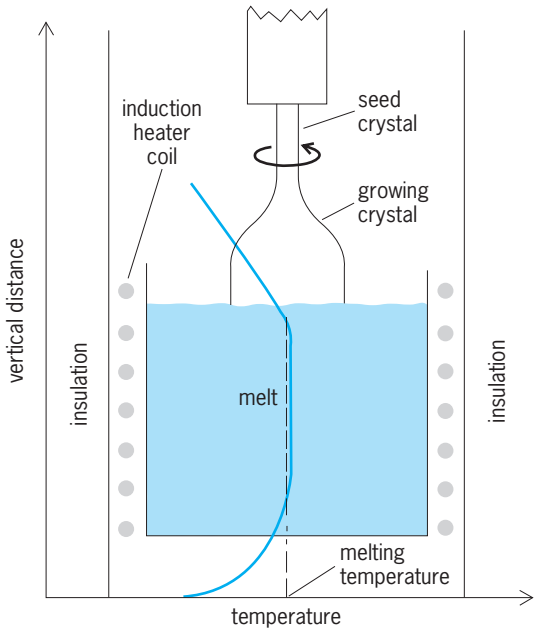

Fig. 5. Czochralski crystal growth and temperature distribution.

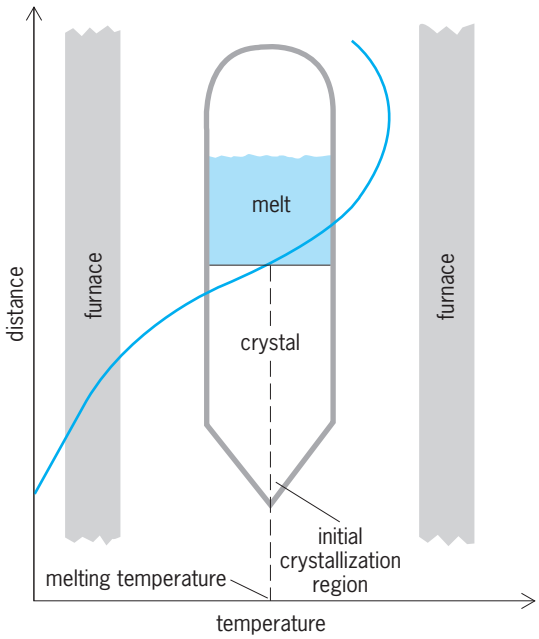

Fig. 6. Vertical Bridgman crystal and temperature distribution.

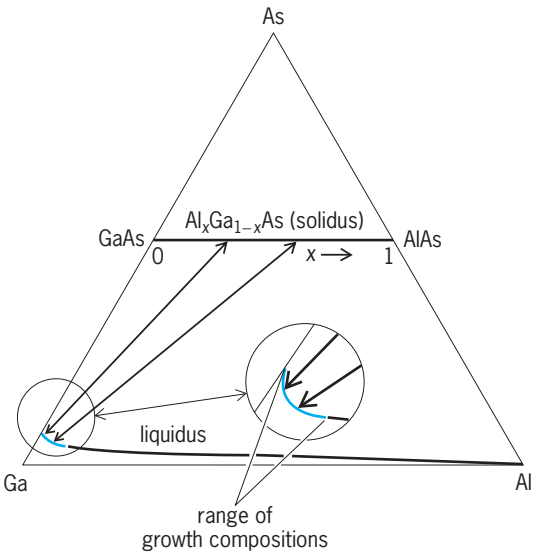

Fig. 7. An isotherm in the aluminum-gallium-arsenic (Al-Ga-As) ternary phase diagram. Each corner represents a pure element and each point within the diagram a unique composition, all at a given temperature. Each composition on the liquidus is in equilibrium with one composition on the solidus. The tie lines (lines with arrowheads) show two such sets of points.

crystal consisting of the entire volume of the melt can result. For some variants of this technique, the liquid is rich in the less volatile component of the solid compound. Growth is driven by maintaining the partial pressure of the more volatile component as required to maintain the liquidus composition. *See* ZONE REFINING.

Solvent methods. Other solution growth methods utilize a solvent in which the crystal components may have extensive solubility, but which itself is not usually appreciably soluble in the crystal. Often, small seed crystals are used to initiate growth which is maintained by subsequent cooling of a saturated solution or diffusion of components through a temperature gradient from a source nutrient. Garnets and other magnetic oxides are often grown from molten salts; quartz crystals are routinely grown from aqueous solution at high temperature and pressure. *See* CRYSTAL WHISKERS; FERRIMAGNETIC GARNETS; SINGLE CRYSTAL.

Epitaxial growth. The evolution of methods for the growth of very thin but very high-quality epitaxial layers has resulted largely from the need for such layers of semiconductors and magnetic garnets.

Liquid-phase epitaxy. The technique most closely related to the methods used for bulk crystal growth is liquid-phase epitaxy. For a typical binary semiconductor, growth is done onto a substrate single-crystal wafer from a solution rich in the component with the lowest partial pressure (Fig. 4). For a binary compound, the grown layers may differ only in impurity concentrations to modify their electrical characteristics. More often, multilayered structures with layers differing in major component composition but having the same crystal structure and lattice parameter are required. The simplest example of liquid-phase epitaxy with major composition changes in layers is the growth of layers of aluminum gallium arsenide ($Al_xGa_{1-x}As$; $1>x>0$) on a gallium arsenide (GaAs) substrate. In this example, for all x, the solid solution $Al_xGa_{1-x}As$ has essentially the same lattice parameter, so that this is a naturally lattice-matched system. An isotherm of the ternary phase diagram for this system is shown in **Fig. 7**. A typical range of growth compositions can yield layers with very different solid compositions.

Growth by liquid-phase epitaxy is done in an apparatus in which the substrate wafer is sequentially brought into contact with solutions that are at the desired compositions and may be supersaturated or cooled to achieve growth. For crystalline solid solutions other than $Al_xGa_{1-x}As$, very precise control over solution compositions is required to achieve a lattice match. Typically, structures grown by liquid-phase epitaxy have four to six layers ranging widely in composition and having thickness from 10^{-7} to 10^{-6} m.

Molecular-beam epitaxy. The desirability of highly reproducible growth and even thinner epitaxial layers of 13–15 compounds on large wafer areas has led to the development of molecular-beam epitaxy and several forms of chemical-vapor deposition. Molecular-beam epitaxy is an ultrahigh-vacuum technique in which beams of atoms or molecules of the constituent elements of the crystal provided by heated effusion ovens, impinge upon a heated substrate crystal (**Fig. 8**). It has been used for epitaxial layers as thin as 0.5 nanometer (**Fig. 9**). Molecular-beam epitaxy has also been used for group 12–16 compounds and for silicon. *See* ARTIFICIALLY LAYERED STRUCTURES; MOLECULAR BEAMS.

The ability to achieve epitaxy of compounds under vacuum conditions requires that it be possible to obtain a nominally stoichiometric crystal from

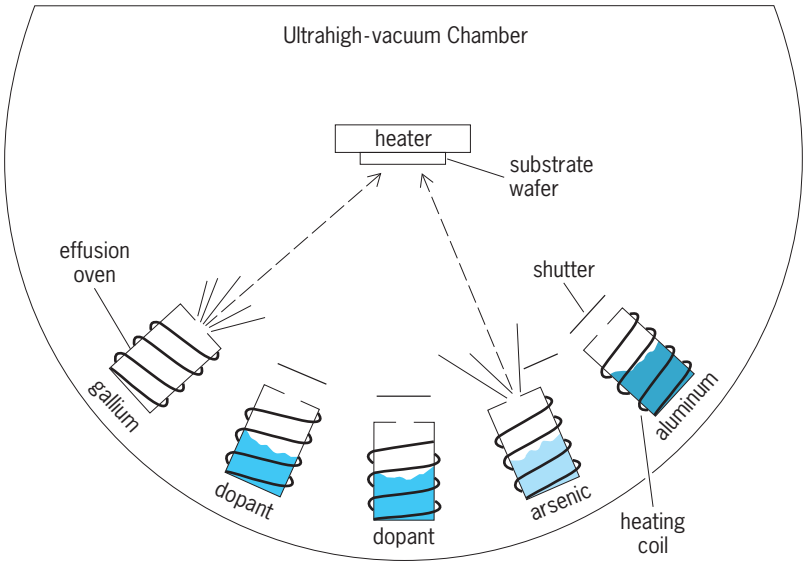

Fig. 8. Molecular-beam epitaxy apparatus.

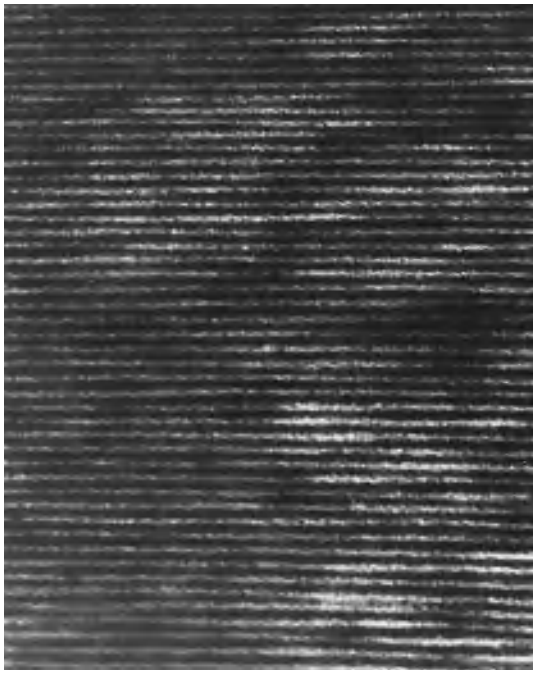

Fig. 9. Cross section of a gallium arsenide–aluminum arsenide (GaAs–AlAs) single crystal with individual gallium arsenide and aluminum arsenide layers each about 5 × 10^{-10} m thick.

impingement of a flux of component atoms and molecules whose net composition is far from that at stoichiometry. It is also required that surface adsorption and mobility of the impinging components be sufficient to ensure that atoms finally reside in their most stable lattice sites. How this can happen is best illustrated with gallium arsenide. Growth of gallium arsenide is achieved by impinging Ga atoms and As_2 or As_4 molecules onto a heated GaAs wafer. The Ga atoms essentially all stick to the surface because of the relatively low Ga vapor pressure at the growth temperature of 900–1300°F (500–700°C). Only those As atoms that interact with a Ga atom stay on the surface, so that the growing layer is stoichiometric provided the As-to-Ga ratio striking the surface is somewhat greater than unity. Crystalline solid solutions such as $Al_xGa_{1-x}As$ are obtained by adding aluminum (Al) to the beams with an Al-containing effusion oven (Fig. 8). Shutters are used to block the beams and permit changes in the beam composition for multilayer growth. *See* GALLIUM.

Chemical-vapor deposition. Chemical-vapor deposition of epitaxial layers is also often used for semiconductors. Again, the group 13–15 compounds are of major importance. The growth of GaAs will be used for illustration. One of the most common chemical-vapor deposition methods is a hot-wall procedure in which the growing layer is generated by reaction (1). The

$$As_4 + 4GaCl + 2H_2 \rightleftharpoons 4GaAs + 4HCl \quad (1)$$

As_4 is generated either by the thermal decomposition of arsine, AsH_3, as in reaction (2), or by reaction (3). (GaCl) is generated by reaction (4). The

$$4AsH_3 \longrightarrow 6H_2 + As_4 \quad (2)$$

$$4AsCl_3 + 6H_2 \longrightarrow As_4 + 12HCl \quad (3)$$

$$2HCl + 2Ga \longrightarrow 2GaCl + H_2 \quad (4)$$

reactants are introduced into a heated fused-silica reaction tube with a carrier gas, usually molecular hydrogen (H_2) or nitrogen (N_2). A temperature gradient is maintained so that the substrate seed is downstream of the decomposition region for reactions (1), (2), and (3), and is at a temperature that favors growth of GaAs by reaction (1). Crystalline solid solutions of a variety of 13–15 compounds can be grown by this process. Layers of $Al_xGa_{1-x}As$ are not generally grown this way because of thermodynamic limitations on composition control, and the very high reactivity of Al-containing vapor species with most hot-wall reactor materials.

Metal organic chemical-vapor deposition. An alternate process is the metal organic (MOCVD) process, in which metal organic compounds are decomposed near the surface of the heated substrate wafer. This method depends upon the fact that a very large variety of MR_x compounds (where M is a metal and R is an organic radical) such as $Ga(CH_3)_3$, $As(CH_3)_4$, and $Zn(CH_3)_2$ are available. The organic radical may be a methyl, ethyl, or more complex group. Generally, a compound is available as a liquid with a convenient vapor pressure so that it can be transported by saturating a stream of H_2. Most often, group 15 elements are transported as the hydrides. Thus, typically, indium phosphide (InP) is grown by reaction (5).

$$In(C_2H_5)_3 + PH_3 \longrightarrow InP + 3C_2H_6 \quad (5)$$

Growth by MOCVD is most often done in a cold-wall chamber containing an induction-heated platform for the substrate wafer. Decomposition of the reactants apparently occurs in a boundary-layer region near the crystal surface. The MOCVD technique has proved highly successful for the growth of GaAs, $Al_xGa_{1-x}As$, InP, and $Ga_xIn_{1-x}P_yAs_{1-y}$, among others, and its versatility for multilayer growth of very thin, very high-quality layers is comparable to that of molecular-beam epitaxy. *See* SEMICONDUCTOR; SEMICONDUCTOR HETEROSTRUCTURES. Morton B. Panish

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Crystal optics

The study of the propagation of light, and associated phenomena, in crystalline solids. The propagation of light in crystals can actually be so complicated that not all the different phenomena are yet completely understood, and not all theoretically predicted phenomena have been demonstrated experimentally.

For a simple cubic crystal the atomic arrangement is such that in each direction through the crystal the crystal presents the same optical appearance. The atoms in anisotropic crystals are closer together in some planes through the material than in others. In anisotropic crystals the optical characteristics are different in different directions. In classical physics the progress of an electromagnetic wave through a material involves the periodic displacement of electrons. In anisotropic substances the forces resisting these displacements depend on the displacement direction. Thus the velocity of a light wave is different in different directions and for different states of polarization. The absorption of the wave may also be different in different directions. *See* DICHROISM; TRICHROISM.

In an isotropic medium the light from a point source spreads out in a spherical shell. The light from a point source embedded in an anisotropic crystal spreads out in two wave surfaces, one of which travels at a faster rate than the other. The polarization of the light varies from point to point over each wave surface, and in any particular direction from the source the polarization of the two surfaces is opposite. The characteristics of these surfaces is opposite. The characteristics of these surfaces can be determined experimentally by making measurements on a given crystal.

For a transparent crystal the theoretical optical behavior is well enough understood so that only a few measurements need to be made in order to predict the behavior of a light beam passing through the crystal in any direction. It is important to remember that the velocity through a crystal is not a function of position in the crystal but only of the direction through the lattice. For information closely related to the ensuing discussion. *See* CRYSTAL STRUCTURE; POLARIZED LIGHT; REFRACTION OF WAVES.

Index ellipsoid. In the most general case of a transparent anisotropic medium, the dielectric constant is different along each of three orthogonal axes. This means that when the light vector is oriented along each direction, the velocity of light is different. One method for calculating the behavior of a transparent anisotropic material is through the use of the index ellipsoid, also called the reciprocal ellipsoid, optical indicatrix, or ellipsoid of wave normals. This is the surface obtained by plotting the value of the refractive index in each principal direction for a linearly polarized light vector lying in that direction (**Fig. 1**). The different indices of refraction, or wave velocities associated with a given propagation direction, are then given by sections through the origin of the coordinates in which the index ellipsoid is drawn. These sections are ellipses, and the major and minor axes of the ellipse represent the fast and slow axes for light proceeding along the normal to the plane of the ellipse. The length of the axes represents the refractive indices for the fast and slow wave, respectively. The most asymmetric type of ellipsoid has three unequal axes. It is a general rule in crystallography that no property of a crystal will have less symmetry than the class in which the crystal belongs. In other words, if a property of the crystal had lower symmetry, the crystal would belong in a different class. Accordingly, there are many crystals which, for example, have four- or sixfold rotation symmetry about an axis, and for these the index ellipsoid cannot have three unequal axes but is an ellipsoid of revolution. In such a crystal, light will be propagated along this axis as though the crystal were isotropic, and the velocity of propagation will be independent of the state of polarization. The section of the index ellipsoid at right angles to this direction is a circle. Such crystals are called uniaxial and the mathematics of their optical behavior is relatively straightforward. *See* CRYSTALLOGRAPHY.

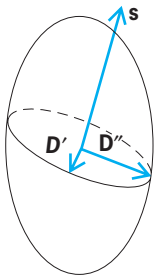

Fig. 1. Index ellipsoid, showing construction of directions of vibrations of D vectors belonging to a wave normal s. (*After M. Born and E. Wolf, Principles of Optics, 6th ed., Pergamon Press, 1980*)

Ray ellipsoid. The normal to a plane wavefront moves with the phase velocity. The Huygens wavelet, which is the light moving out from a point disturbance, will propagate with a ray velocity. Just as the index ellipsoid can be used to compute the phase or wave velocity, so can a ray ellipsoid be used to calculate the ray velocity. The length of the axes of this ellipsoid is given by the velocity of the linearly polarized ray whose electric vector lies in the axis direction. *See* PHASE VELOCITY.

The ray ellipsoid in the general case for anisotropic crystal is given by Eq. (1), where α, β, and γ are the

$$\alpha^2 x^2 + \beta^2 y^2 + \gamma^2 z^2 = 1 \tag{1}$$

three principal indices of refraction and where the velocity of light in a vacuum is taken to be unity. From this ellipsoid the ray velocity surfaces or Huygens wavelets can be calculated as just described. These surfaces are of the fourth degree and are given by

Eq. (2). In the uniaxial case $\alpha = \beta$ and Eq. (2) factors

$$\left(x^2+y^2+z^2\right)\left(\frac{x^2}{\alpha^2}+\frac{y^2}{\beta^2}+\frac{z^2}{\gamma^2}\right)$$
$$-\frac{1}{\alpha^2}\left(\frac{1}{\beta^2}+\frac{1}{\gamma^2}\right)x^2-\frac{1}{\beta^2}\left(\frac{1}{\gamma^2}+\frac{1}{\alpha^2}\right)\gamma^2$$
$$-\frac{1}{\gamma^2}\left(\frac{1}{\alpha^2}+\frac{1}{\beta^2}\right)z^2+\frac{1}{\alpha^2\beta^2\gamma^2}=0 \quad (2)$$

into Eqs. (3). These are the equations of a sphere and

$$x^2+y^2+z^2=\frac{1}{\alpha^2}$$
$$\gamma^2(x^2+y^2)+\alpha^2z^2=1 \quad (3)$$

an ellipsoid. The z axis of the ellipsoid is the optic axis of the crystal.

The refraction of a light ray on passing through the surface of an anisotropic uniaxial crystal can be calculated with Huygens wavelets in the same manner as in an isotropic material. For the ellipsoidal wavelet this results in an optical behavior which is completely different from that normally associated with refraction. The ray associated with this behavior is termed the extraordinary ray. At a crystal surface where the optic axis is inclined at an angle, a ray of unpolarized light incident normally on the surface is split into two beams: the ordinary ray, which proceeds through the surface without deviation; and the extraordinary ray, which is deviated by an angle determined by a line drawn from the center of one of the Huygens ellipsoidal wavelets to the point at which the ellipsoid is tangent to a line parallel to the surface. The construction is shown in **Fig. 2**. The two beams are oppositely linearly polarized. When the incident beam is inclined at an angle ϕ to the normal, the ordinary ray is deviated by an amount determined by Snell's law of refraction, the extraordinary ray by an amount which can be determined in a manner similar to that of the normal incidence case already described. The plane wavefront in the crystal is first found by constructing Huygens wavelets as shown in **Fig. 3**. The line from the center of the wavelet to the point of tangency gives the ray direction and velocity.

The relationship between the normal to the wavefront and the ray direction can be calculated algebraically in a relatively straightforward fashion. The extraordinary wave surface, or Huygens wavelet, is given by Eq. (3), which can be rewritten as Eq. (4),

$$\epsilon^2(x^2+y^2)+\omega^2z^2=1 \quad (4)$$

where ϵ is the extraordinary index of refraction and ω the ordinary index. A line from the center of this ellipsoid to a point (x_1, y_1, z_1) on the surface gives the velocity of a ray in this direction. The wave normal corresponding to this ray is found by dropping a perpendicular from the center of the ellipsoid to the plane tangent at the point (x_1, y_1, z_1). For simplicity consider the point in the plane $y = 0$. The tangent

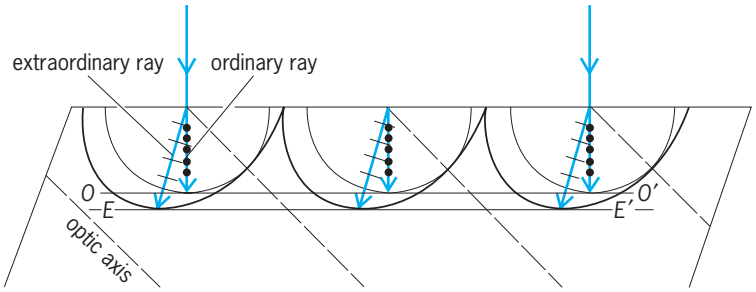

Fig. 2. Huygens construction for a plane wave incident normally on transparent calcite. If one proceeds to find the common tangents to the secondary wavelets shown, the two plane waves labeled *OO′* and *EE′* are obtained. (*After F. A. Jenkins and H. E. White, Fundamentals of Optics, 4th ed., McGraw-Hill, 1976*)

at the point x_1z_1 is given by Eq. (5). The slope of the

$$\epsilon^2xx_1+\omega^2zz_1=1 \quad (5)$$

normal to this line is given by Eq. (6). The tangent

$$\frac{z}{x}=\frac{\omega^2z_1}{\epsilon^2x_1} \quad (6)$$

of the angle between the optic axis and the wave normal is the reciprocal of this number, as shown in Eq. (7), where ψ is the angle between the ray and the

$$\tan\varphi=\frac{\epsilon^2x_1}{\omega^2z_1}=\frac{\epsilon^2}{\omega^2}\tan\psi \quad (7)$$

optic axis. The difference between these two angles, π, can be calculated from Eq. (8). This quantity is a maximum when Eq. (9) holds.

$$\tan\tau=\frac{\tan\varphi-\tan\psi}{1+\tan\varphi\tan\psi} \quad (8)$$

$$\tan\varphi=\pm\frac{\epsilon}{\omega} \quad (9)$$

One of the first doubly refracting crystals to be discovered was a transparent variety of calcite called Iceland spar. This uniaxial crystal cleaves into slabs in which the optic axis makes an angle of 45° with one pair of surfaces. An object in contact with or a few inches from such a slab will thus appear to be doubled. If the slab is rotated about a normal to

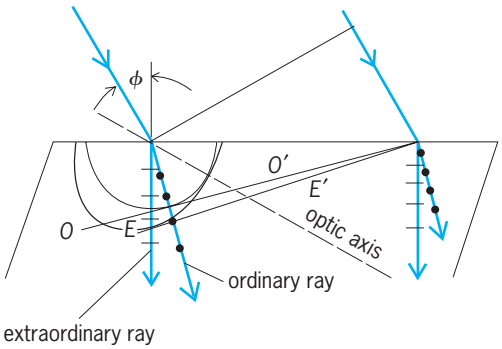

Fig. 3. Huygens construction when the optic axis lies in the plane of incidence. (*After F. A. Jenkins and H. E. White, Fundamentals of Optics, 4th ed., McGraw-Hill, 1976*)

the surface, one image rotates about the other. For the sodium D lines at 589 nanometers, the indices for calcite are given by Eqs. (10). From these, the

$$\epsilon = 1.486 \qquad \omega = 1.659 \tag{10}$$

maximum angle τ_{max} between the wave normal and the ray direction is computed to be 6°16′. The wave normal at this value makes an angle of 41°52′ with the optic axis. This is about equal to the angle which the axis makes with the surface in a cleaved slab. Accordingly, the natural rhomb gives nearly the extreme departure of the ray direction from the wave normal.

Interference in polarized light. One of the most interesting properties of plates of crystals is their appearance in convergent light between pairs of linear, circular, or elliptical polarizers. An examination of crystals in this fashion offers a means of rapid identification. It can be done with extremely small crystals by the use of a microscope in which the illuminating and viewing optical systems are equipped with polarizers. Such a polarizing microscope is a common tool for the mineralogist and the organic chemist.

In convergent light the retardation through a birefringent plate is different for each direction. The slow and fast axes are also inclined at a different angle for each direction. Between crossed linear polarizers the plate will appear opaque at those angles for which the retardation is an integral number of waves. The locus of such points will be a series of curves which represent the characteristic interference pattern of the material and the angle at which the plate is cut. In order to calculate the interference pattern, it is necessary to know the two indices associated with different angles of plane wave propagation through the plate. This can be computed from the index ellipsoid. *See* INTERFERENCE OF WAVES.

Ordinary and extraordinary indices. For the uniaxial crystal there is one linear polarization direction for which the wave velocity is always the same. The wave propagated in this direction is called the ordinary wave. This can be seen directly by inspection of the index ellipsoid. Since it is an ellipsoid of revolution, each plane passing through the center will produce an ellipse which has one axis equal to the axis of the ellipsoid. The direction of polarization will always be at right angles to the plane of incidence and the refractive index will be constant. This constant index is called the ordinary index. The extraordinary index will be given by the other axis of the ellipse and will depend on the propagation direction. When the direction is along the axis of the ellipsoid, the extraordinary index will equal the ordinary index.

From the equation of the ellipsoid, Eq. (11), one

$$\frac{x^2}{\omega^2} + \frac{y^2}{\omega^2} + \frac{z^2}{\epsilon^2} = 1 \tag{11}$$

can derive the expression for the ellipse and in turn Eq. (12) for the extraordinary index n_e as a function

$$n_e = \frac{\omega\epsilon}{(\epsilon^2 \cos^2 r + \omega^2 \sin^2 r)^{1/2}} \tag{12}$$

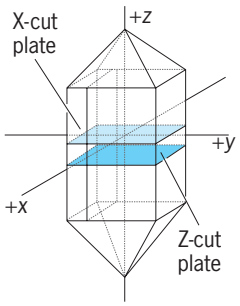

Fig. 4. X-cut and Z-cut plates in $NH_4H_2PO_4$ crystal.

of propagation direction. Since the ellipsoid is symmetrical it is necessary to define this direction only with respect to the ellipsoid axis. Here r is the angle in the material between the normal to the wavefront and the axis of the ellipsoid, n_e is the extraordinary index associated with this direction, ω is the ordinary index, and ϵ is the maximum value of the extraordinary index (usually referred to simply as the extraordinary index).

A slab cut normal to the optic axis is termed a Z-cut or C-cut plate (**Fig. 4**). When such a plate is placed between crossed linear polarizers, such as Nicol prisms, it gives a pattern in monochromatic light shown in **Fig. 5**. The explanation of this pattern is obtained from the equations given for the indices of refraction. To a first approximation the retardation for light of wavelength λ passing through the plate at an angle r can be written as Eq. (13), where d is

$$\Gamma = \frac{(n_e - \omega)d}{\lambda \cos r} \tag{13}$$

the thickness of the plate. The axis of the equivalent retardation plate at an angle r will be in a plane containing the optic axis of the plate and the direction of light propagation. Wherever Γ is a whole number

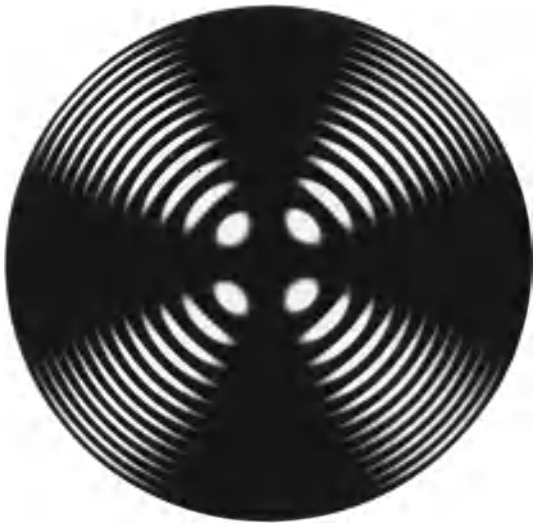

Fig. 5. Interference figure from fluorspar that has been cut perpendicular to the optic axis and placed between crossed Nicol prisms. (*After M. Born and E. Wolf, Principles of Optics, 6th ed., Pergamon Press, 1980*)

of waves, the light leaving the plate will have the same polarization as the incident light. The ordinary index is constant. The extraordinary index is a function of r alone, as seen from Eq. (12). Accordingly, the locus of direction of constant whole wave retardation will be a series of cones. If a uniform white light background is observed through such a plate, a series of rings of constant whole wave retardation will be seen. Since the retardation is a function of wavelength, the rings will appear colored. The innermost ring will have the least amount of color. The outermost ring will begin to overlap and disappear as the blue end of the spectrum for one ring covers the red end of its neighbor. The crystal plate has no effect along the axes of the polarizers since here the light is polarized along the axis of the equivalent retardation plate. The family of circles is thus bisected by a dark cross. When the crystal is mounted between like circular polarizers, the dark cross is not present and the center of the system of rings is clear. During World War II crystal plates were used in this fashion as gunsights. The rings appear at infinity even when the plate is close to the eye. Lateral movement of the plate causes no angular motion of the ring system. The crystal plate could be rigidly fastened to a gun mount and adjusted so as to show at all times the direction in space in which the gun was pointing.

Uniaxial crystal plates in which the axis lies in the plane of the plate are termed X-cut or Y-cut. When such plates are observed between crossed polarizers, a pattern of hyperbolas is observed.

When crystal plates are combined in series, the patterns become much more complex. In addition to the fringes resulting from the individual plates, a system of so-called moiré fringes appears. *See* MOIRÉ PATTERN.

Negative and positive crystals. In calcite the extraordinary wave travels faster than the ordinary wave. Calcite and other materials in which this occurs are termed negative crystals. In positive crystals the extraordinary wave travels slower than the ordinary wave. In the identification of uniaxial minerals one of the steps is the determination of sign. This is most easily demonstrated in a section cut perpendicular to the optic axis. Between crossed linear polarizers the pattern in convergent light, as already mentioned, is a series of concentric circles which are bisected by a dark cross. When a quarter-wave plate is inserted between one polarizer and the crystal, with its axis at 45° to the polarizing axis, the dark rings are displaced outward or inward in alternate quadrants. If the rings are displaced outward along the slow axis of the quarter-wave plate, the crystal is positive. If the rings are displaced inward along the axis, then the crystal is negative.

If a Z-cut plate of a positive uniaxial crystal is put in series with a similar plate of a negative crystal, it is possible to cancel the birefringence so that the combination appears isotropic.

Biaxial crystals. In crystals of low symmetry the index ellipsoid has three unequal axes. These crystals are termed biaxial and have two directions along which the wave velocity is independent of the polarization direction. These correspond to the two sections of the ellipsoid which are circular. These sections are inclined at equal angles to the major axes of the ellipsoid, and their normals lie in a plane containing the major and intermediate axes. In convergent light between polarizers, these crystals show a pattern which is quite different from that which appears with uniaxial crystals. A plate cut normal to the major axis of the index ellipsoid has a pattern of a series of lemniscates and ovals. The directions corresponding to the optic axes appear as black spots between crossed circular polarizers. The interference pattern between crossed linear polarizers is shown in **Fig. 6**.

Fig. 6. Interference figure from Brazil topaz. (*After M. Born and E. Wolf, Principles of Optics, 6th ed., Pergamon Press, 1980*)

Angle between optic axes. One of the quantities used to describe a biaxial crystal or to identify a biaxial mineral is the angle between the optic axes. This can be calculated directly from the equation for the index ellipsoid, Eq. (14), where α, β, and γ

$$\frac{x^2}{\alpha^2} + \frac{y^2}{\beta^2} + \frac{z^2}{\gamma^2} = 1 \tag{14}$$

are the three indices of refraction of the material. The circular sections of the ellipsoid must have the intermediate index as a radius. If the relative sizes of the indices are so related that $\alpha > \beta > \gamma$, the circular sections will have β as a radius and the normal to these sections will lie in the xz plane. The section of the ellipsoid cut by this plane will be the ellipse of Eq. (15). The radius of

$$\frac{x^2}{\alpha^2} + \frac{z^2}{\gamma^2} = 1 \tag{15}$$

length β will intersect this ellipse at a point x_1z_1 where Eq. (16) holds. The solution of these two

$$x_1^2 + z_1^2 = \beta^2 \tag{16}$$

equations gives for the points x_1z_1 Eqs. (17). These

$$x_1 = \pm\sqrt{\frac{\alpha^2(\gamma^2-\beta^2)}{\gamma^2-\alpha^2}}$$
$$z_1 = \pm\sqrt{\frac{\gamma^2(\alpha^2-\beta^2)}{\alpha^2-\gamma^2}} \qquad (17)$$

points and the origin define the lines in the xz plane which are also in the planes of the circular sections of the ellipsoid. Perpendiculars to these lines will define the optic axes. The angle Ω between the axes and the z direction will be given by Eq. (18).

$$\tan^2\Omega = \frac{{z_1}^2}{{x_1}^2} = \frac{(1/\alpha^2)-(1/\beta^2)}{(1/\beta^2)-(1/\gamma^2)} \qquad (18)$$

The polarization direction for light passing through a biaxial crystal is computed in the same way as for a uniaxial crystal. The section of the index normal to the propagation direction is ordinarily an ellipse. The directions of the axes of the ellipse represent the vibration direction and the lengths of the half axes represent the indices of refraction. One polarization direction will lie in a plane which bisects the angle made by the plane containing the propagation direction and one optic axis and the plane containing the propagation direction and the other optic axis. The other polarization direction will be at right angles to the first. When the two optic axes coincide, this situation reduces to that which was demonstrated earlier for the case of uniaxial crystals.

Conical refraction. In biaxial crystals there occurs a set of phenomena which have long been a classical example of the theoretical prediction of a physical characteristic before its experimental discovery. These are the phenomena of internal and external conical refraction. They were predicted theoretically in 1832 by William Hamilton and experimentally demonstrated in 1833 by H. Lloyd.

As shown earlier, the Huygens wavelets in a biaxial crystal consist of two surfaces. One of these has its major axis at right angles to the major axis of the other. The two thus intersect. In making the geometrical construction to determine the ray direction, two directions are found where the two wavefronts coincide and the points of tangency on the two ellipsoids are multiple. In fact, the plane which represents the wavefront is found to be tangent to a circle which lies partly on one surface and partly on the other. The directions in which the wavefronts coincide are the optic axes. A ray incident on the surface of a biaxial crystal in such a direction that the wavefront propagates along an axis will split into a family of rays which lie on a cone. If the crystal is a plane parallel slab, these rays will be refracted at the second surface and transmitted as a hollow cylinder parallel to the original ray direction. Similarly, if a ray is incident on the surface of a biaxial crystal plate at such an angle that it passes along the axes of equal ray velocity, it will leave the plate as a family of rays lying on the surface of a cone. The first of these phenomena is internal conical refraction; the second is external conical refraction. Equation (19) gives the

$$\tan\psi = \frac{\beta}{\sqrt{\alpha\gamma}}\sqrt{(\beta-\alpha)(\gamma-\beta)} \qquad (19)$$

half angle ψ of the external cone. Bruce H. Billings

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Crystal structure

The arrangement of atoms, ions, or molecules in a crystal. Crystals are defined as solids having, in all three dimensions of space, a regular repeating internal unit of structure. This definition, which dates from early studies of crystals, expresses very concisely what crystals are.

The interior of crystals has been studied by the use of x-rays, which excite signals from the atoms that inhabit the crystal. The signals are of different strengths and depend on the electron density distribution about atomic cores. Thus, there is no probe analogous to a constant light that illuminates a landscape uniformly. Light atoms give weaker signals and hydrogen is invisible to x-rays. However, the mutual atomic arrangements that are called crystal structures can be derived once the chemical formulas and physical densities of solids are known, based on the knowledge that atomic positions are not arbitrary but are dictated by crystal symmetry, and that the diffraction signals received are the result of systematic constructive interference between the scatterers within the regularly repeating internal unit of pattern.

This article discusses fundamental concepts, the dimensionality dependence of crystal structure determination, the structures found for elements and chemically simple compounds, and the conclusions drawn. The relationship between low coordination and structural variability is illustrated, as is a noncrystallographic packing. For related information *See* CRYSTALLOGRAPHY; POLYMORPHISM (CRYSTALLOGRAPHY); X-RAY CRYSTALLOGRAPHY; X-RAY DIFFRACTION.

Fundamental Concepts

Crystals should be imagined as three-dimensional solids. They have fundamental units of structure—atoms, ions, or molecules—mutually arranged to form a geometrically regular unit of pattern, or motif, that continuously repeats itself by translations in three dimensions. Crystals are thus defined in terms of space, population, and mutual arrangement. Crystal space is represented as an indefinitely extended lattice of periodically repeating points. The periodicity of the lattice is defined by the lengths and mutual orientations of three lattice vectors that enclose the pattern. Population is defined as the total number

and kind of fundamental units of structure that form the pattern.

Ordering and states of matter. It is the chemical ordering to make a pattern, and the geometric regularity with which that pattern is repeated in space, that distinguishes among the three classical states of matter—gas, liquid, and solid—and also distinguishes between the crystals and noncrystalline solids. Gases are freely moving atoms or molecules that have no interaction that maintains a contact between them. They are totally disordered. Liquids of various viscosities are essentially incompressible. This means that there are fundamental units of structure that maintain a contact distance between them as they move freely in space. This distance is called the short-range spatial order of the liquid. Short-range order is similar to the order in a bag of marbles. *See* GAS; LIQUID.

Glasses are similar to liquids but have rigid structures because contact between fundamental units is maintained. Their geometric regularity of repetition extends to a few diameters of the fundamental unit. This is an intermediate range of spatial order, larger than that of a liquid. *See* AMORPHOUS SOLID; GLASS.

The order and periodicity of crystals must extend to about 100 nanometers in all three dimensions of space to give the sharply defined diffraction signals required for mapping structural details by x-rays. Intermediate states of order are seen in liquid crystals, which have long molecules as fundamental units of structure. These are arranged with their lengths parallel to each other, but without periodicity, in the nematic state. In the smectic state there is orientation in equally spaced planes but no sideways periodicity, like traffic moving freely on a multilane highway. *See* LIQUID CRYSTALS.

Requirement for three-dimensional order. The necessity of meeting the order and periodicity requirements in three dimensions for exact mapping of crystal structure is well illustrated by elemental carbon. In the graphite structure (**Fig. 1**) the atoms are arranged in six-member, planar, hexagonal rings, which form two-dimensional extensive and periodic layers by sharing their edges. Every carbon atom is trigonally bonded to three others in the same sheet, with separations of 0.142 nm. Normal to this sheet, the layer spacing is 0.342 nm, but this does not define a period. There are no binding forces between carbon atoms in different sheets, except for the weak van der Waals attractive forces; the layers therefore slide over each other like cards in a deck. There is therefore no lattice periodicity in the third dimension. The structure is fundamentally two-dimensional, and was more difficult to define than the three-dimensional, tetrahedrally bonded diamond structure, where the carbon-carbon separation is 0.154 nm. There are no sharp diffraction signals from which the interlayer spacing could be determined precisely because, lattice periodicity being absent in this direction, there is no fixed carbon-carbon separation. *See* GRAPHITE.

Arrangement of atoms and molecules. **Figure 2** illustrates the lattice symmetry and the fundamental

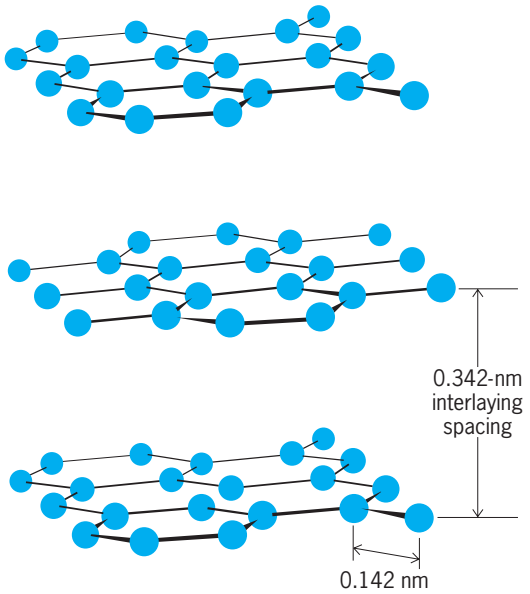

Fig. 1. Crystal structure of graphite.

unit of pattern and symmetry operations of a crystal structure that generate clones of the pattern in different orientations. The link between a crystal structure and either physical properties or biological function requires examination of the mutual arrangement of all the atoms and molecules, the distances that separate atom centers, and the distribution of empty space.

Figure 2 shows the crystal structure of anthracene ($C_{14}H_{10}$), an organic dye with blue fluorescence. The crystal symmetry is monoclinic with three unequal lattice vectors ($a = 0.856$ nm, $b = 0.604$ nm, $c = 1.116$ nm), and one angle, $\beta \neq 90^\circ$ ($\beta = 124.7^\circ$). There are 14 carbon atoms, arranged as three fused hexagonal rings in each unit of pattern and two units of pattern in each unit cell: eight at the cell vertices, each shared by eight cells (8/8 = 1), and two halves within each cell. The long axes of the molecules are parallel, inclined at about 9° from the c axis of the crystal. The 10 hydrogen atoms, invisible to x-rays

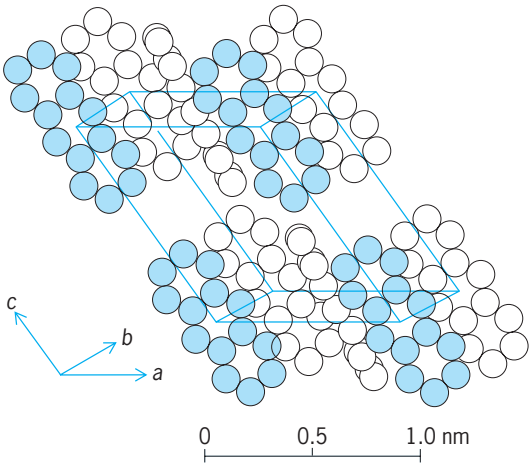

Fig. 2. Structure of anthracene ($C_{14}H_{10}$).

but seen by neutron diffraction since they have a magnetic moment, symmetrically surround each unit of pattern. *See* NEUTRON DIFFRACTION.

The cohesive forces that cause these nonpolar molecules to come together and pack themselves into a crystal are explained by quantum mechanics as instantaneously induced dipole moments of one molecule on another, due to the noncoincidence of the centers of charge of an atom's nucleus and the electron cloud. Their average over time is zero, so that they are not observed as a net dipole moment. They provide the small cohesive energy of the crystal and the rapidly varying, minute locations of lines of communication within the crystal. *See* CHEMICAL BONDING; COHESION (PHYSICS); INTERMOLECULAR FORCES.

Crystal imperfection. The realities of crystal space are, in general, not as they are found in perfect diamonds, nor as they are modeled to be, even in crystals that appear to be freely formed, as in the case of crystallization from supersaturated solutions or from a melt. If scaled up, these crystals would win no prizes in crafting; not in quiltmaking for accurate replication of details within a unit of pattern; nor in wall building for precise translations of identical unit-cell bricks; nor in architecture for the exact symmetries of their extensive designs. The growth process is characterized by constraints and turbulences, and by the dynamic interaction between the substance that is crystallizing and the medium that provides the components of growth and the growth environment. The process is reflected within the structures formed as an assemblage of atoms is collected and made relatively immobile by releasing the energy known as the heat of crystallization. The resulting crystal lattices resemble a mosaic of slightly misaligned adjacent regions. This is actually fortunate for research in x-ray crystallography. Perfect alignment would result in subtraction of energy by interference with the primary beam, due to a 180° phase reversal of the reflected beam (primary extinction). Internally diffracted beams would also be attenuated by internal reflection from regions above them (secondary extinction). *See* CRYSTAL GROWTH.

Detailed structure and advanced probes. Each of the spatially misaligned mosaic blocks of a single crystal is assumed to maintain lattice periodicity within it. This assumption is confirmed by the sharp diffraction patterns observed. There are some "wrong" atoms, vacant lattice sites, trapped gas atoms, and so forth, and the atomic occupants jiggle about while also vibrating cooperatively and synchronously in complex internal modes of motion. Intricate patterns of electron exchange are enacted, and systematic changes in spin orientations can occur for an atom with a magnetic moment. *See* CRYSTAL DEFECTS; EXCHANGE INTERACTION; LATTICE VIBRATIONS.

Details like these are important for understanding the relationships between structure determination on the atomic and molecular levels and the cooperative behavior that determines bulk properties and functions. They are explored by methods other than conventional x-ray crystallography. For example, the power and tunability of synchrotron x-radiation makes it possible to discriminate against a specific atom by not exciting it to generate a signal. In effect, it is possible to sequentially "light up" different atoms and resolve specific details by parts. Intricate structures with very large atomic populations thereby yield maps with details resolved. *See* SYNCHROTRON RADIATION.

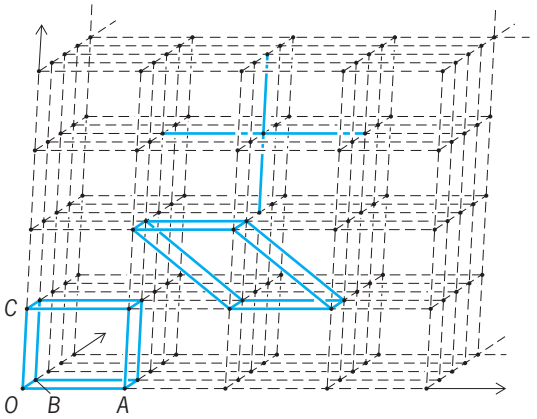

Fig. 3. A space lattice, two possible unit cells, and the environment of a point.

Space Lattices and Symmetry Operations

Figure 3 shows a rectangular space lattice with two possible cells outlined. These have the same cell volumes but different symmetries. Since crystallographic unit cells are completely defined by three lattice vectors, the crystal symmetry referenced to this lattice can be no higher than orthorhombic: $a \neq b \neq c$ ($OA \neq OB \neq OC$), and all angles equal to 90°. This and a possible monoclinic cell, with the same vectors a and b (OA and OB) and one angle not equal to 90°, are outlined. If the OAB plane is rotated and the vector a (OA) is extended to terminate at the next lattice point, then all angles differ from 90° and the crystal symmetry represented becomes triclinic. The mutual arrangement and atom coordinates of the cell population must be such that the environment, seen from every point of the space lattice, remains the same.

Screw axes. These combine the rotation of an ordinary symmetry axis with a translation parallel to it and equal to a fraction of the unit distance in this direction. **Figure 4** illustrates such an operation, which is symbolically denoted 3_1 and 3_2. The translation is respectively 1/3 and 2/3. The helices are added to help the visualization, and it is seen that they are respectively right- and left-handed. The projection on a plane perpendicular to the axis shows that the relationship about the axis remains in spite of the displacement. A similar type of arrangement can be considered around the other symmetry axes, and the following possibilities arise: 2_1, 3_1, 3_2, 4_1, 4_2, 4_3, 6_1, 6_2, 6_3, 6_4, and 6_5.

If screw axes are present in crystals, it is clear that the displacements involved are of the order of a few tenths of nanometer and that they cannot be

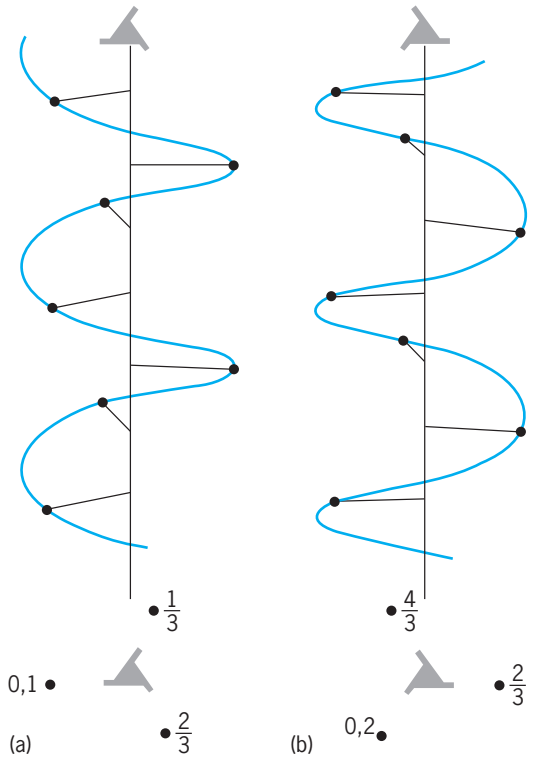

Fig. 4. Screw axes (*a*) 3_1 and (*b*) 3_2. Lower parts of figure are projections on a plane perpendicular to the axis. Numbers indicate heights of points above that plane.

distinguished macroscopically from ordinary symmetry axes. The same is true for glide mirror planes, which combine the mirror image with a translation parallel to the mirror plane over a distance that is half the unit distance in the glide direction.

The handedness of screw axes is a very important feature of many biological and mineral structures. They come in enantiomorphic pairs, with one member of the pair, the left-handed one, preferred, in the biological molecules found on Earth. In concentrated sugar solutions this screw symmetry operation is apparently maintained in some fashion since different directions for rotation of the plane of polarized light are observed. *See* OPTICAL ACTIVITY; STEREOCHEMISTRY.

The helices of deoxyribonucleic acid (DNA) and of quartz are alike in that they contain what is essentially a structural backbone that can both flex and change partners with molecules that are not in the structure while maintaining structural continuity. In quartz, this is reflected by the direction of negative expansion when heated. The helical chain of $(SiO_4)^{4-}$ vertex-linked tetrahedra changes shape as two linked tetrahedra rotate about their common oxygen vertex. The overall configurational change is like that of a squashed mattress coil when it is sat upon. *See* DEOXYRIBONUCLEIC ACID (DNA); SILICA MINERALS.

Space groups. These are indefinitely extended arrays of symmetry elements disposed on a space lattice. A space group acts as a three-dimensional kaleidoscope: An object submitted to its symmetry operations is multiplied and periodically repeated in such a way that it generates a number of interpenetrating identical space lattices. The fact that 230 space groups are possible means, of course, 230 kinds of periodic arrangement of objects in space. When only two dimensions are considered, 17 space groups are possible.

Space groups are denoted by the Hermann-Mauguin notation preceded by a letter indicating the Bravais lattice on which it is based. For example, P $2_12_12_1$ is an orthorhombic space group; the cell is primitive and three mutually perpendicular screw axes are the symmetry elements. J. D. H. Donnay and D. Harker have shown that it is possible to deduce the space group from a detailed study of the external morphology of crystals.

Common Structures

A discussion of some common structural systems found in metals, crystalline compounds, and minerals is given below.

Metals. In general, metallic structures are relatively simple, characterized by a close packing and a high degree of symmetry. Manganese, gallium, mercury, and one form of tungsten are exceptions. Metallic elements situated in the subgroups of the periodic table gradually lose their metallic character and simple structure as the number of the subgroup increases. A characteristic of metallic structures is the frequent occurrence of allotropic forms; that is, the same metal can have two or more different structures which are most frequently stable in a different temperature range.

The forces which link the atoms together in metallic crystals are nondirectional. This means that each atom tends to surround itself by as many others as possible. This results in a close packing, similar to that of spheres of equal radius, and yields three distinct systems: close-packed (face-centered) cubic, hexagonal close-packed, and body-centered cubic.

Close packing. For spheres of equal radius, close packing is interesting to consider in detail with respect to metal structures. Close packing is a way of arranging spheres of equal radius in such a manner that the volume of the interstices between the spheres is minimal. The problem has an infinity of solutions. The manner in which the spheres can be most closely packed in a plane *A* is shown in **Fig. 5**. Each sphere is in contact with six others; the centers form a regular pattern of equilateral triangles. The cavities between the spheres are numbered. A second, similar plane can be positioned in such a way that its spheres rest in the cavities 1, 3, 5 between those of the layer *A*. The new layer, *B*, has an arrangment similar to that of *A* but is shifted with respect to *A*. Two possibilities exist for adding a third layer. Its spheres can be put exactly above those of layer *A* (an assembly *ABA* is then formed), or they may come above the interstices 2, 4, 6. In the latter case the new layer is shifted with respect to *A* and *B* and is called *C*. For each further layer two possibilities exist, and any sequence such as *ABCBABCACBA* . . . in which two successive layers have not the same denomination is a solution of the problem. In the vast majority of

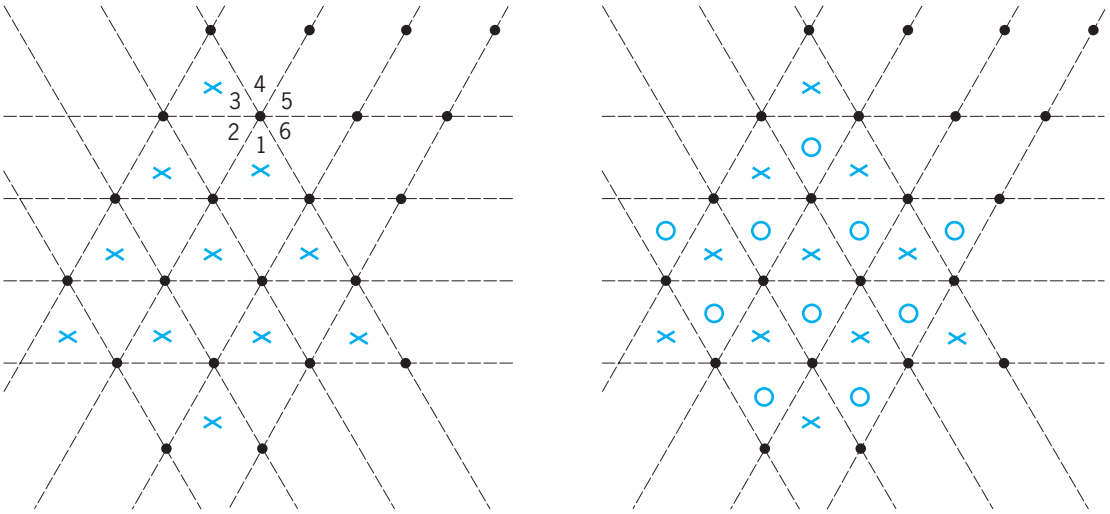

Fig. 5. Close packing of spheres of equal radius in a plane. The centers for the *A*, *B*, and *C* layers are indicated by a dot, a cross, and an open circle, respectively. Cavities between spheres are numbered.

cases, periodic assemblies with a very short repeat distance occur.

Face-centered cubic structure. This utilizes close packing characterized by the regular repetition of the sequence *ABC*. The centers of the spheres form a cubic lattice, as shown in **Fig. 6***a*. This form contains four sets of planes. The densely packed planes of the type *A, B, C* are perpendicular to the threefold axis and can therefore be written as $\{111\}$. Since these are the closest packed planes of the structure, d_{111} is greater than any other d_{hkl} of the lattice. The densest rows in these planes are $\langle 110 \rangle$.

It is relatively easy to calculate the percentage of the volume occupied by the spheres. The unit cell has a volume a^3 and contains four spheres of radius R, their volume being $16\pi R^3/3$. The spheres touch each other along the face diagonal (Fig. 6*a*); $a\sqrt{2}$ is therefore equal to $4R$, or $R = a\sqrt{2}/4$. Substitution gives for the volume of the spheres $\pi\sqrt{2a^3/6}$, which is 73% of the volume of the cube. It is clear that the same percentage of the unit volume is filled in all other close-packed assemblies, that is, assemblies corresponding to other alternations of the planes *A, B, C*.

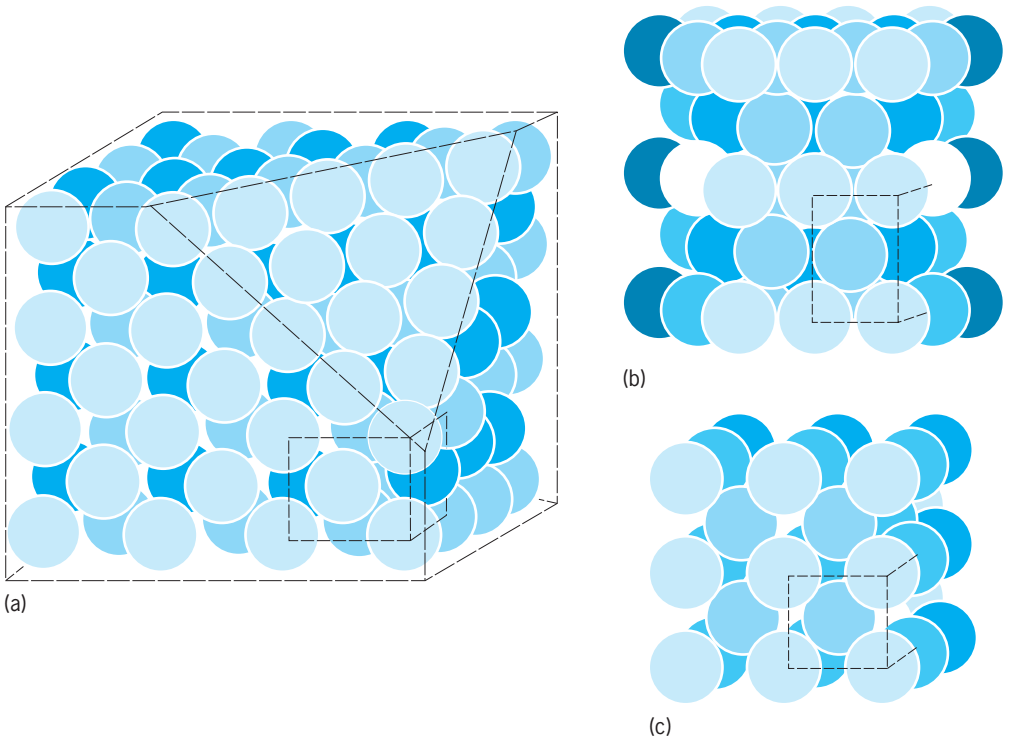

Fig. 6. Close packing of spheres in space. (*a*) Cubic close packing (face-centered cubic). One set of close-packed 111″ planes (*A*, *B*, or *C*) is shown. (*b*) Hexagonal close packing. (*c*) Body-centered cubic arrangement.

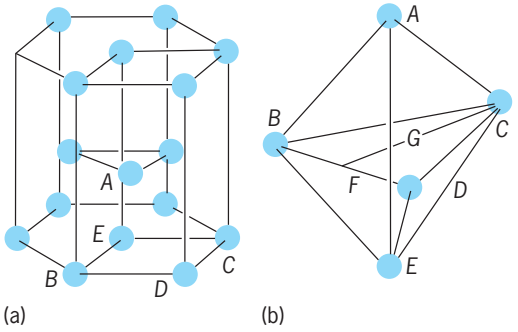

Fig. 7. Hexagonal close-packed structure. (*a*) Three unit cells, showing how the hexagonal axis results. One of the cells is fully outlined. (*b*) Calculation of the ratio *c*/*a*. Distance *AE* is equal to height of cell.

Hexagonal close-packed structure. This is a close packing characterized by the regular alternation of two layers, or *ABAB* . . . . The assembly has hexagonal symmetry (**Fig. 7**). Six spheres are at the corners of an orthogonal parallelepiped having a parallelogram as its base; another atom has as coordinates (1/3, 1/3, 1/2). The ratio *c*/*a* is easily calculated. The length $BD = a$, the edge of the unit cell. The height of the cell is $AE = 2AG = c$, defined in Eqs. (1) and (2).

$$AG = \sqrt{a^2 - \left(\frac{2}{3}\frac{a\sqrt{3}}{2}\right)^2} = a\sqrt{2/3} \qquad (1)$$

$$c = 2AG = a\sqrt{8/3} \qquad c/a = \sqrt{8/3} = 1.633 \qquad (2)$$

This latter value is important, for it permits determination of how closely an actual hexagonal structure approaches ideal close packing.

Body-centered cubic structure. This is an assembly of spheres in which each one is in contact with eight others (Fig. 6*c*).

The spheres of radius *R* touch each other along the diagonal of the cube, so that measuring from the centers of the two corner cubes, the length of the cube diagonal is 4*R*. The length of the diagonal is also equal to $a\sqrt{3}$, if *a* is the length of the cube edge, and thus $R = a\sqrt{3/4}$. The unit cell contains two spheres (one in the center and 1/8 in each corner) so that the total volume of the spheres in each cube is $8\pi R^3/3$. Substituting $R = a\sqrt{3/4}$ and dividing by a^3, the total volume of the cube, gives the percentage of filled space as 67%. Thus the structure is less dense than the two preceding cases.

The closest-packed planes are {110}; this form contains six planes. They are, however, not as dense as the *A, B, C* planes considered in the preceding structures. The densest rows have the four ⟨111⟩ directions.

Tabulation of structures. The structures of various metals are listed in **Table 1**, in which the abbreviations fcc, hcp, and bcc, respectively, stand for face-centered cubic, hexagonal close-packed, and body-centered cubic. The structures listed hcp are only roughly so; only magnesium has a *c*/*a* ratio (equal to 1.62) that is very nearly equal to the ratio (1.63) calculated above for the ideal hcp structure. For cadmium and zinc the ratios are respectively 1.89 and 1.86, which are significantly larger than 1.63. Strictly speaking, each zinc or cadmium atom has therefore not twelve nearest neighbors but only six. This departure from the ideal case for subgroup metals follows a general empirical rule, formulated by W. Hume-Rothery and known as the 8-*N* rule. It states that a subgroup metal has a structure in which each atom has 8-*N* nearest neighbors, *N* being the number of the subgroup.

Table 2 compares metallic (elemental) and ionic radii of the same elements. The first and second elements of their respective periods are listed. Both radii are determined from families of compounds. Metallic radii are always larger—for the light metals, very much larger—than the ionic radii. Thus, although the metallic structures are close-packed arrangements, the atoms obviously do not touch each other as they do in the sphere-packing models.

Crystalline compounds. Simple crystal structures are usually named after the compounds in which they were first discovered (diamond or zinc sulfide, cesium chloride, sodium chloride, and calcium fluoride). Many compounds of the types A^+X^- and $A^{2+}X_2^-$ have such structures. They are highly

TABLE 1. Metal structures

Metal	Modification	Stability range	Structure
Cadmium	α	To melting point	hcp
Chromium	α	To melting point	bcc
	β	Electrolytic form	Hexagonal
Cobalt	α	To 788°F (420°C)	hcp mixed with fcc
	β	788°F (420°C) to melting point	fcc
Gold	α	To melting point	fcc
Iron	α	To 1668°F (909°C)	bcc
	γ	1668–2557°F (909–1403°C)	fcc
	δ	2557°F (1403°C) to melting point	bcc
Magnesium	α	To melting point	Nearly hcp
Nickel	α	To melting point	fcc
	β	Electrolytic form	Hexagonal
Silver	α	To melting point	fcc
Zinc	α	To melting point	hcp
Zirconium	α	To 1584°F (862°C)	hcp
	β	1584°F (862°C) to melting point	bcc

TABLE 2. Metallic and ionic radii in crystals

Period	Element	Atomic number Z	Metallic structure: Structure type	Metallic structure: Radius, nm	Ionic structure: Radius, nm	Ionic structure: Charge
First short	Lithium (Li)	3	bcc	0.152	0.07	1^+
	Beryllium (Be)	4	hcp	0.111	0.035	2^+
Second short	Sodium (Na)	11	bcc	0.186	0.10	1^+
	Magnesium (Mg)	12	hcp	0.160	0.07	2^+
First long	Potassium (K)	19	bcc	0.231	0.133	1^+
	Calcium (Ca)	20	fcc	0.197	0.100	2^+
Second long	Rubidium (Rb)	37	bcc	0.240	0.150	1^+
	Strontium (Sr)	38	fcc	0.210	0.110	2^+
Third long	Cesium (Cs)	55	bcc	0.266	0.170	1^+
	Barium (Ba)	56	bcc	0.217	0.135	2^+

symmetrical, the unit cell is cubic, and the atoms or ions are disposed at the corners of the unit cell and at points having coordinates that are combinations of 0, 1, 1/2, or 1/4.

Sodium chloride structure. This is an arrangement in which each positive ion is surrounded by six negative ions, and vice versa. The arrangement is expressed by stating that the coordination is 6/6. The centers of the positive and the negative ions each form a face-centered cubic lattice. They are shifted one with respect to the other over a distance $a/2$, where a is the repeat distance (**Fig. 8***a*). Systematic study of the dimensions of the unit cells of compounds having this structure has revealed that:

1. Each ion can be assigned a definite radius. A positive ion is smaller than the corresponding atom and a negative ion is larger. Figure 8*b* shows the effective sizes of ions in sodium chloride.

2. Each ion tends to surround itself by as many others as possible of the opposite sign because the binding forces are nondirectional.

On this basis the structure is determined by two factors, a geometrical factor involving the size of the two ions which behave in first approximation as hard spheres, and an energetical one involving electrical neutrality in the smallest possible volume. In the ideal case all ions will touch each other; therefore, if r_A and r_X are the radii of the ions, $4r_X = a\sqrt{2}$ and $2(r_A + r_X) = a$. Expressing a as a function of r_X gives $r_A/r_X = \sqrt{2} - 1 = 0.41$. When r_A/r_X becomes smaller than 0.41, the positive and negative ions are no longer in contact and the structure becomes unstable. When r_A/r_X is greater than 0.41, the positive ions are no longer in contact, but ions of different sign still touch each other. The structure is stable up to $r_A/r_X = 0.73$, which occurs in the cesium chloride structure. *See* IONIC CRYSTALS.

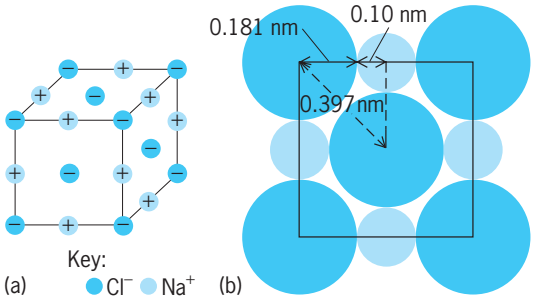

Fig. 8. Structure of sodium chloride. (*a*) Cubic unit cell. (*b*) Basal plane of the unit cell. Effective sizes of ions are shown.

Cesium chloride structure. This is characterized by a coordination 8/8 (**Fig. 9***a*). Each of the centers of the positive and negative ions forms a primitive cubic lattice; the centers are mutually shifted over a distance $a\sqrt{3/2}$. The stability condition for this structure can be calculated as in the preceding case. Contact of the ions of opposite sign here is along the cube diagonal (Fig. 9*b*).

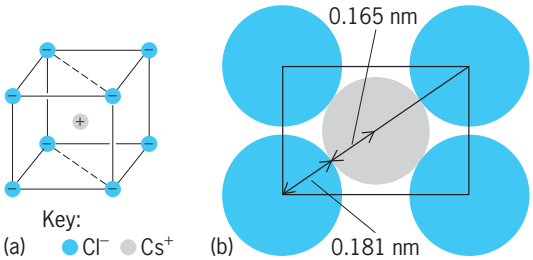

Fig. 9. Structure of cesium chloride. (*a*) Cubic unit cell. (*b*) Section of the cell through a diagonal plane, showing the close packing of Cs^+ and Cl^- ions, and effective sizes of ions.

Diamond structure. In this arrangement each atom is in the center of a tetrahedron formed by its nearest neighbors. The 4-coordination follows from the well-known bonds of the carbon atoms. This structure is illustrated in **Fig. 10***a*. The atoms are at the corners of the unit cell, the centers of the faces, and at four points having as coordinates (1/4, 1/4, 1/4), (3/4, 3/4, 1/4), (3/4, 1/4, 3/4), (1/4, 3/4, 3/4). The atoms can be divided into two groups, each forming a face-centered cubic lattice; the mutual shift is $a\sqrt{3/4}$.

Zinc blende structure. This structure, shown in Fig. 10*b*, has coordination 4/4 and is basically similar to the diamond structure. Each zinc atom (small circles in Fig. 10*b*) is in the center of a tetrahedron formed by sulfur atoms (large circles), and vice versa. The zinc atoms form a face-centered cubic lattice, as do the sulfur atoms.

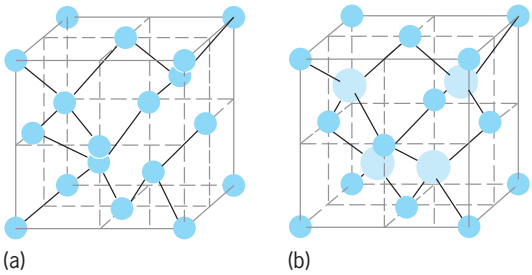

Fig. 10. Tetrahedral crystal compound structures. (*a*) Diamond. (*b*) Zinc blende.

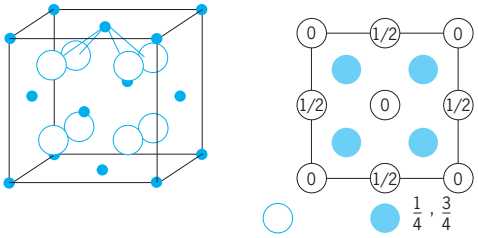

Fig. 11. Calcium fluoride structure.

Calcium fluoride structure. **Figure 11** shows the calcium fluoride structure. If the unit cell is divided into eight equal cubelets, calcium ions are situated at corners and centers of the faces of the cell. The fluorine ions are at the centers of the eight cubelets. There exist three interpenetrating face-centered cubic lattices, one formed by the calcium ion and two by the fluorine ions, the mutual shifts being (0, 0, 0), (1/4, 1/4, 1/4), (3/4, 3/4, 3/4).

Minerals. Silica (SiO_2) is the most abundant material in the Earth's crust. It is surpassed only by carbon in variety of architectural structures. As in carbon, these structures also have widely ranging, dynamic properties. **Figure 12** indicates schematically how this abundance of structures can evolve from the 4:2 coordination of silica where each silicon atom has four tetrahedrally disposed oxygen atoms about it. Figure 12*a* represents an isolated tetrahedron with unsatisfied charge about each oxygen atom. Joining tetrahedra to make a single chain (Fig. 12*b*) in effect subtracts oxygen atoms and lowers the charge. Joining two such chains (Fig. 12*c*) extends the structure and also reduces the charge since another oxygen is subtracted. It is readily seen how the single chain of Fig. 12*b* could be wrapped as a helix, thereby extending the charged regions into the third dimension. *See* SILICATE MINERALS.

Figure 13 shows a polyhedral structural model of the zeolite faujasite. The composition is $NaCa_{0.5}(Al_2Si_5O_{14}) \cdot 10H_2O$. Every vertex of the truncated octahedra with hexagonal faces is the location of a tetrahedrally coordinated silicon or aluminum atom. The structure continues to fill space indefinitely. However, while model space is filled according to lattice requirements, physical space is quite empty. This is a charged framework structure in which mobile species, such as water (H_2O) molecules and cations, can freely flow through the large pipes that are an integral part of the structure. Cation exchange occurs with this flow. *See* MOLECULAR SIEVE; ZEOLITE.

Structures and Periodicity

The requirement that structures have three-dimensional periodicity that can be referenced to a Bravais lattice of mathematical points with three lattice vectors, that every point of the Bravais lattice have the same environment about it, and therefore that identical unit cells repeat periodically by translations in all three directions, is a requirement for accurate crystal structure determination by x-rays. As was demonstrated by the example of graphite, it is not a requirement imposed upon physical reality. Compositionally modulated structures exist; indeed, they are to be expected. As growth occurs, local composition changes, the heat of crystallization is released, and growing surfaces become available as templates for fresh deposition. Nor does the fact that crystals, as defined, give sharp x-ray diffraction patterns imply that all diffraction patterns come from crystals, as defined.

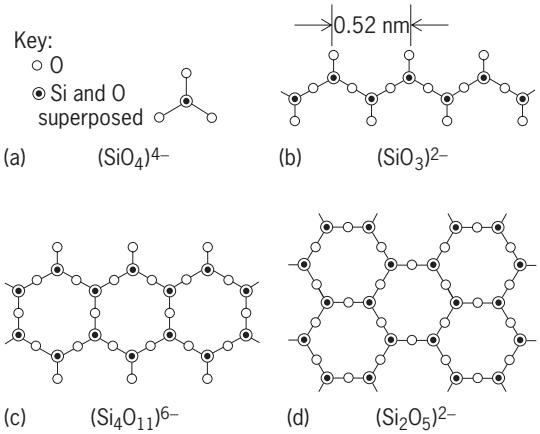

Fig. 12. Structures of silicates. (*a*) Fundamental SiO_4 tetrahedron. (*b*) Linear chain of tetrahedra. (*c*) Band formed by two linked linear chains. (*d*) Plane of linked tetrahedra.

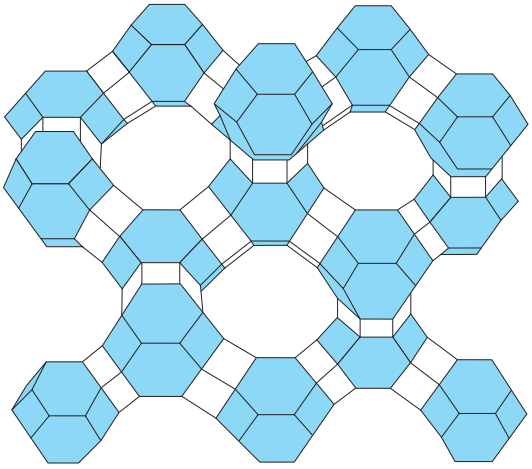

Fig. 13. Crystal structure of faujasite. The silicon and aluminum atoms are situated at the apices of the truncated octahedra, which are joined to form a three-dimensional framework.

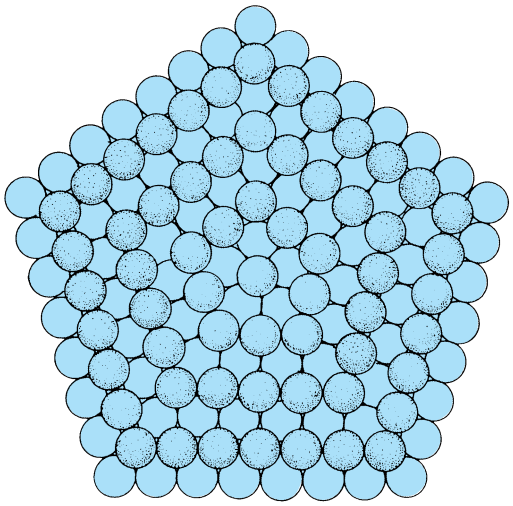

Fig. 14. Sphere packing with fivefold symmetry. (*After A. F. Wells, Models in Structural Inorganic Chemistry, Oxford University Press, 1970*)

Finally, the random packing of spheres in contact, resembling a bag of marbles, is very dense, while close-packed arrangements need not be physically dense. **Figure 14** shows how a dense-sphere packing, with one-dimensional periodicity only, along a fivefold symmetry axis, will fill space. A two-layer construction is shown in which the first layer forms concentric pentagons with odd numbers of spheres on each side, and a second layer is added in which the corresponding numbers of spheres are even. The compact two-layer constructions extend indefinitely in two dimensions and can be stacked to form an infinite packing along the fivefold axis. The density is 0.724, which is denser than the body-centered cubic close packing of spheres. Doris Evans

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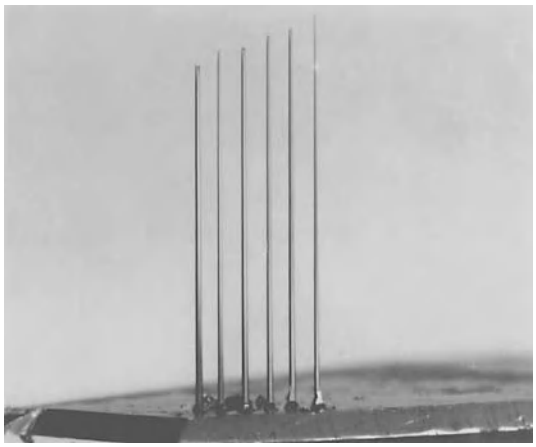
Crystal whiskers

Single crystals that have grown in filamentary form. Such filamentary growths have been known for centuries and can form by a variety of mechanisms. Some grow from their base: either these are extruded to relieve pressure in the base material or they grow as a result of a chemical reaction at the base. In both cases, the growth occurs at a singularity in the base material. Other crystal whiskers grow at their tip, into a supersaturated medium, and the filamentary growth results from a strong anisotropy in the growth rate. *See* SINGLE CRYSTAL; SUPERSATURATION.

The growth of extruded metal whiskers was studied in the 1950s, when it was discovered that tin whiskers growing from solder were causing electrical short circuits. These whiskers were typically a few micrometers in diameter and a few millimeters in length. This shorting problem was solved by adding more lead to the solder. Great interest in the whiskers developed after it was discovered that the strength exhibited by some whiskers approached that expected theoretically for perfect crystals. This great strength results from the internal and surface perfection of the whiskers, whereas the strength of most materials is limited by defects. The interest in the high strength of the whiskers centered on the possibility of using them in composites to increase the strength of more ductile matrix materials. Fibers of silica, boron, and carbon, which are much easier to fabricate in large quantity than whiskers, also exhibit similarly high strengths, and are now used in composites. *See* COMPOSITE MATERIAL; CRYSTAL DEFECTS.

The controlled growth of whiskers of a wide variety of materials is possible by means of the vapor-liquid-solid mechanism (see **illus.**). Here, a liquid



Crystal whiskers grown by the vapor-liquid-solid process. (*AT&T Bell Laboratories*)

droplet containing impurities rides on the tip of the whisker, providing a preferential site for condensation. Many electronic materials are prone to whisker formation by the vapor-liquid-solid process during chemical vapor deposition, so that control of the deposition to eliminate whisker formation is essential. Whiskers of many substances have also been grown in supersaturated solutions, but the mechanism of growth is not understood in detail. *See* CRYSTAL GROWTH. Kenneth A. Jackson; Richard S. Wagner

Bibliography. S. S. Brenner, Growth and properties of "whiskers," *Science*, 128:569–575, 1958; C. Kittel, *Introduction to Solid State Physics*, 7th ed., 1996; A. P. Levitt, *Whisker Technology*, 1970.

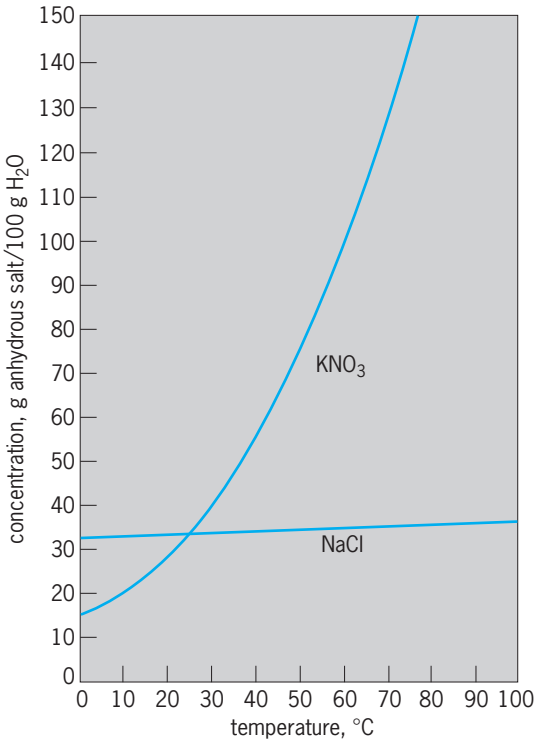
Crystallization

The formation of a solid from a solution, melt, vapor, or a different solid phase. Crystallization from solution is an important industrial operation because

of the large number of materials marketed as crystalline particles. Fractional crystallization is one of the most widely used methods of separating and purifying chemicals. This article discusses crystallization of substances from solutions and melts. *See* GLASS; PHASE EQUILIBRIUM; POLYMER; SINGLE CRYSTAL; SOLUTION; SUBLIMATION.

Solutions. In order for crystals to nucleate and grow, the solution must be supersaturated; that is, the solute must be present in solution at a concentration above its solubility. Different methods may be used for creating a supersaturated solution from one which is initially undersaturated. The possible methods depend on how the solubility varies with temperature. Two examples of solubility behavior are shown in the **illustration**. Either evaporation of water or cooling may be used to crystallize potassium nitrate (KNO_3), while only evaporation would be effective for NaCl. An alternative is to add a solvent such as ethanol which greatly lowers the solubility of the salt, or to add a reactant which produces an insoluble product. This causes a rapid crystallization perhaps more properly known as precipitation. *See* PRECIPITATION (CHEMISTRY).

Nucleation. The formation of new crystals is called nucleation. At the extremely high supersaturations produced by addition of a reactant or a lower-solubility solvent, this nucleation may take place in the bulk of the solution in the absence of any solid surface. This is known as homogeneous nucleation. At more moderate supersaturations new crystals form on solid particles or surfaces already present in the solution (dust, motes, nucleation catalysts, and so on). This is called heterogeneous nucleation.



Temperature-solubility curves for two salts. $°F = (°C \times 1.8) + 32$.

When solutions are well agitated, nucleation is primarily secondary, that is, from crystals already present. Probably this is usually due to minute pieces breaking off the crystals by impact with other crystals, with the impeller, or with the walls of the vessel. *See* NUCLEATION.

Crystal size. Generally, large crystals are considered more desirable than small crystals, probably in the belief that they are more pure. However, crystals sometimes trap (occlude) more solvent as inclusions when they grow larger. Thus there may be an optimal size. The size distribution of crystals is influenced primarily by the supersaturation, the amount of agitation, and the growth time. Generally, the nucleation rate increases faster with increasing supersaturation than does the growth rate. Thus, lower supersaturations, gentle stirring, and long times usually favor large crystals. Low supersaturations require slow evaporation and cooling rates.

Crystal habit. Often the habit (shape) of crystals is also an important commercial characteristic. In growth from solutions, crystals usually display facets along well-defined crystallographic planes, determined by growth kinetics rather than by equilibrium considerations. The slowest-growing facets are the ones that survive and are seen. While habit depends somewhat on the supersaturation during growth, very dramatic changes in habit are usually brought about by additives to the solution. Strong habit modifiers are usually incorporated into the crystal, sometimes preferentially. Thus the additive may finally be an impurity in the crystals.

Fractional crystallization. In fractional crystallization it is desired to separate several solutes present in the same solution. This is generally done by picking crystallization temperatures and solvents such that only one solute is supersaturated and crystallizes out. By changing conditions, other solutes may be crystallized subsequently. Occasionally the solution may be supersaturated with respect to more than one solute, and yet only one may crystallize out because the others do not nucleate. Preferential nucleation inhibitors for the other solutes or seeding with crystals of the desired solute may be helpful. Since solid solubilities are frequently very small, it is often possible to achieve almost complete separation in one step. But for optimal separation it is necessary to remove the impure mother liquor from the crystals. Rinsing of solution trapped between crystals is more effective for large-faceted crystals. Internally occluded solution cannot be removed by rinsing. Even high temperatures may not burst these inclusions from the crystal. Repeated crystallizations are necessary to achieve desired purities when many inclusions are present or when the solid solubility of other solutes is significant.

Melts. If a solid is melted without adding a solvent, it is called a melt, even though it may be a mixture of many substances. That is the only real distinction between a melt and a solution. Crystallization of a melt is often called solidification, particularly if the process is controlled by heat transfer so as to produce a relatively sharp boundary between the solid

and the melt. It is then possible to slowly solidify the melt and bring about a separation, as indicated by the phase diagram. The melt may be stirred to enhance the separation. The resulting solid is usually cut into sections, and the purest portion is subjected to additional fractional solidifications. Alternatively the melt may be poured off after part of it has solidified. Zone melting was invented to permit multiple fractional solidifications without the necessity for handling between each step. Fractionation by the above techniques appears to be limited to purification of small batches of materials already fairly pure, say above 95%. *See* ZONE REFINING.

Fractional crystallization from the melt is also being used for large-scale commercial separation and purification of organic chemicals. It has also been tested for desalination of water. Rather than imposing a sharp temperature gradient, as in the above processes, the melt is relatively isothermal. Small, discrete crystals are formed and forced to move countercurrent to the melt. At the end from which the crystals exit, all or part of the crystals are melted. This melt then flows countercurrent to the crystals, thereby rinsing them of the adhering mother liquor. *See* CRYSTAL GROWTH; CRYSTAL STRUCTURE; CRYSTALLOGRAPHY.

William R. Wilcox

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Crystallography

The branch of science that deals with the geometric forms of crystals. The field of crystallography originated in the prehistoric observation of geometric forms of minerals, such as a collection of quartz (SiO_2) crystals (**Fig. 1**). These crystals grow as hexagonal pyramids, with growth being terminated by the coming together of external faces to form a pinnacle. Equally beautiful forms can be made by very simple procedures. How to describe, classify, and measure such forms are the first questions of crystallography. Revealing the forces that made them and the activities within them are the modern directions of the field. This article deals with the rigor, scope, and impact of crystallography, its interdisciplinary nature, and the key results and successes that have made it historically continuous and vital. For other aspects of crystallography *see* CRYSTAL STRUCTURE; POLYMORPHISM (CRYSTALLOGRAPHY).

Crystallography is essential to progress in the applied sciences and technology and developments in all materials areas, including metals and alloys, ceramics, glasses, and polymers, as well as drug design. It is equally vital to progress in fundamental physics and chemistry, mineralogy and geology, and computer science, and to understanding of the dynamics and processes of living systems. This vitality and power are reflected in the rapid exploitation and commercial development of its methods and the support given to data files that record and make available its results.

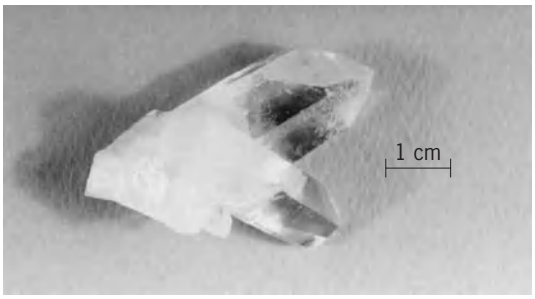


Fig. 1. Natural, intergrown quartz crystals.

Classical Fundamentals

Understanding of crystals with well-developed faces can be obtained with visible light, a simple mechanical device, mathematics, and the imagination as the only tools.

Faces and zones. Slow changes in the composition of the in-feeding substances and in the environment of growth (temperature, pressure, and atmospheric composition) yield many minerals that have large crystals that are optically transparent and chemically well defined, with well-defined shapes. The external morphology reflects growth rates in different directions. These directions remain constant during the course of the growth process, and are represented mathematically as the normals to sets of parallel planes that are imagined as being added on as growth proceeds. The faces that meet and define an edge belong to a zone, a zone being a set of planes that share one common direction, the zone axis. The invariance of interfacial angles, measured by rotation about an axis that is defined by the zone direction, was discovered in the seventeenth century. *See* CRYSTAL GROWTH.

Goniometer. **Figures 2** and **3** show the optical goniometer and the method of tagging and

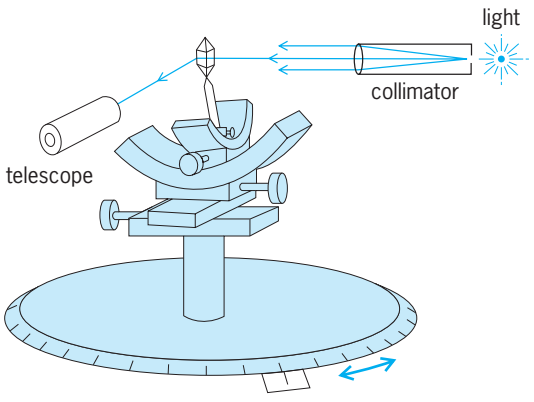


Fig. 2. Reflecting goniometer.

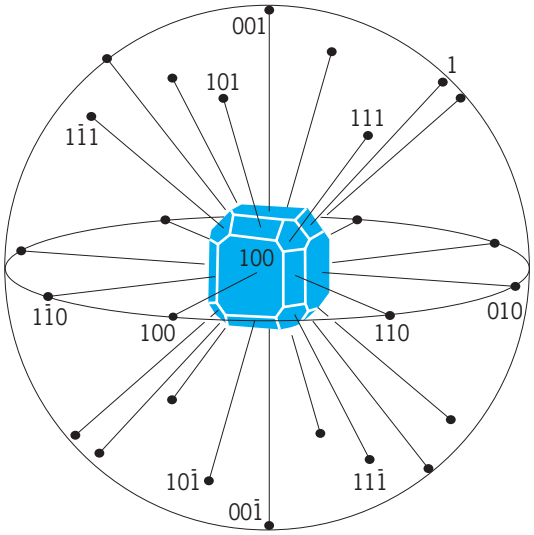


Fig. 3. Spherical projection of normals to crystal faces.

measuring morphological geometries. Translation in two orthogonal directions aligns a face of the crystal so that it reflects light into a telescope when a zone axis is vertical, and the position is recorded. The crystal is then reoriented by rotation about the goniometer axis and two calibrated circular arcs, horizontal or equatorial and vertical or azimuthal, to reflect light from another face into the telescope.

Spherical projection. Interfacial angles are calculated from spherical geometry. Figure 3 illustrates the procedure for a crystal having well-developed faces of which three are mutually perpendicular. The normals to these faces are the natural directions for constructing an orthogonal frame of reference for measurement. The crystal is imagined to be shrunk and placed at the center of a sphere with coordinates (0,0,0). The face normals, labeled [100], [010], and [001], define the directions of an orthogonal reference system. Normals to the same set of planes, but oppositely directed, are labeled [$\bar{1}$00], [0$\bar{1}$0], [00$\bar{1}$]. The reversal of sign indicates that the crystal must be rotated 180° to obtain the same view. Rotation

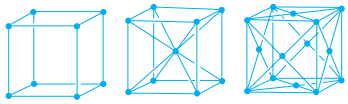
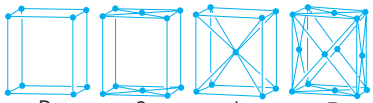
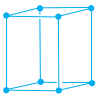
Bravais lattice cells	Axes and interaxial angles	Examples
 Cubic P, Cubic I, Cubic F	Three axes at right angles; all equal: $a = b = c$; $\alpha = \beta = \gamma = 90°$	Copper (Cu), silver (Ag), sodium chloride (NaCl)
 Tetragonal P, Tetragonal I	Three axes at right angles; two equal: $a = b \neq c$; $\alpha = \beta = \gamma = 90°$	White tin (Sn), rutile (TiO_2), β-spodumene ($LiAlSi_2O_6$)
 P, C, I, F Orthorhombic	Three axes at right angles; all unequal: $a \neq b \neq c$; $\alpha = \beta = \gamma = 90°$	Gallium (Ga), perovskite ($CaTiO_3$)
 Monoclinic P, Monoclinic C	Three axes, one pair not at right angles, of any lengths: $a \neq b \neq c$; $\alpha = \gamma = 90° \neq \beta$	Gypsum ($CaSO_4 \cdot 2H_2O$)
 Triclinic P	Three axes not at right angles, of any lengths: $a \neq b \neq c$; $\alpha \neq \beta \neq \gamma \neq 90°$	Potassium chromate (K_2CrO_7)
 Trigonal R (rhombohedral)	Rhombohedral: three axes equally inclined, not at right angles; all equal: $a = b = c$; $\alpha = \beta = \gamma \neq 90°$	Calcite ($CaCO_3$), arsenic (As), bismuth (Bi)
 Trigonal and hexagonal C (or P)	Hexagonal: three equal axes coplanar at 120°, fourth axis at right angles to these: $a_1 = a_2 = a_3 \neq c$; $\alpha = \beta = 90°$, $\gamma = 120°$	Zinc (Zn), cadmium (Cd), quartz (SiO_2) [P]

Fig. 4. The 14 Bravais lattices, derived by centering of the seven crystal classes (P and R) defined by symmetry operators.

about the [001] direction interchanges the positions of [100] and [$\bar{1}$00] faces and their bounding edges. Rotation about [010] turns these faces upside down. Correct designations for group movements and symmetry operations are clearly essential for establishing and maintaining orientation in crystal space. The directions of face normals determine points at which the imagined sphere is pierced. The solid angles between an array of such points, all lying on the same great circle of the sphere, belong to a zone.

Bravais lattices. Optical measurements and stereographic projections established the constancy of interfacial angles, independent of how well developed the faces are. Such properties as the cleavage of large rhombohedral crystals of calcite ($CaCO_3$) into little rhombs suggested that the large crystal could be represented by geometrically identical smaller units stacked together, by translation, to fill space. The 14 lattices of Bravais (**Fig. 4**) enlarged the seven crystal systems of optical mineralogy by adding centering points to them: body (I), face (F), and base (C) centers. The 14 lattices define three-dimensional distributions of mathematical points such that the environments of all points of the lattice are identical. They also define the symmetries of frameworks for constructing mathematical models to represent the observed and measured realities—models made from cells of the smallest volume, but also highest symmetry, that stack together by translation to fill space.

Stacking of model cells does not imply that a crystal grows by stacking identical bricks; a lattice of identically surrounded mathematical points does not imply that any real objects, atoms or molecules, are located at the points; and filling space by translation of identical cells does not imply that the space defined by the cells is filled. Rather, the Bravais lattices are a formalism for representing observed geometries and symmetries of real crystals by three-dimensional lattices of identically surrounded points. **Figure 5** shows a two-dimensional square array of points with imaginary cells outlined. The four different shapes with one lattice point per cell have the same area and one elementary motif per cell (P in Fig. 4). The square with two units of motif, one within it and the four that are common to each of four cells ($4/4 = 1$), has twice the area of the other shapes; it is centered. The lengths of three edges, and the angles between them, represent crystal symmetry with reference to three-dimensional lattices of such points.

Miller indices. The lattices also provide the means to identify imaginary planes within the cell; these are called Miller indices (*h,k,l*). They consist of small whole numbers. For example, each of the six faces of a simple cube, with the origin of a coordinate frame of reference at the cube body center, is normal to one of the reference axes and parallel to the plane defined by the two others. The six faces are indexed as their normals in Fig. 3—(100), ($\bar{1}$00), (010), (0$\bar{1}$0), (001), and (00$\bar{1}$)—to represent a face that intercept the $x,\bar{x}$ axis but not the y and z; the y, $\bar{y}$ axis but not the x and z; and so forth. Hypothetical parallel planes with $^1/_2$ the interplanar spacing are represented as (200), ($\bar{2}$00), (020), and so forth.

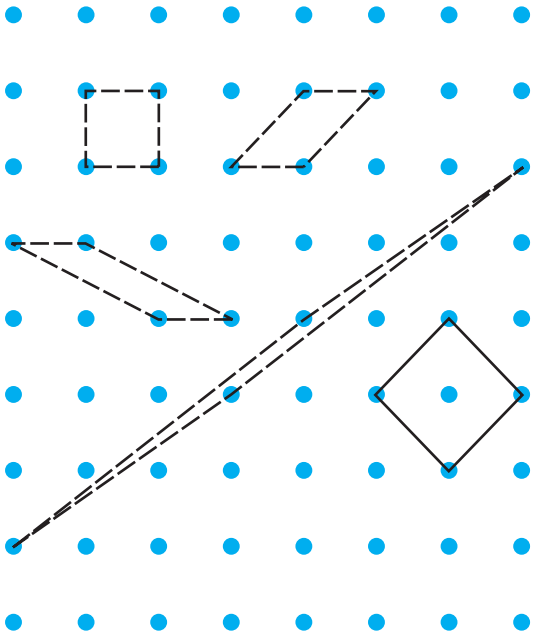


Fig. 5. Two-dimensional lattice of points. All cells fill space by translation; that is, the surroundings of all lattice points are the same. The four cells indicated by broken lines are P, with one unit of pattern and the same areas. The cell indicated by solid lines is C, with two units of pattern and twice the area of the others.

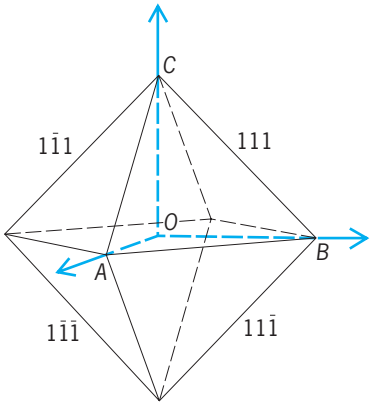


Fig. 6. Indexing of the four front faces of an octahedron.

Figure 6 shows the indexing of faces for an octahedon—the solid obtained by truncating the eight vertices of a cube until all the original faces disappear and a new regular solid with six vertices and eight faces, the dual of the cube with eight vertices and six faces, emerges. The truncation simulates the relative change in growth rates that exposes interior planes as bounding faces, thereby creating the architecture of crystals. Nothing has changed except the faces exposed. The octahedron has full cubic symmetry. The eight bounding faces have intercepts (111), ($\bar{1}$11), (1$\bar{1}$1), (11$\bar{1}$), ($\bar{1}\bar{1}$1), (1$\bar{1}\bar{1}$), ($\bar{1}$1$\bar{1}$), and ($\bar{1}\bar{1}\bar{1}$).

Symmetry operations and space groups. A complete mathematical formalism for modeling an external morphological form, and the symmetry relations between imagined units of structure within it, was in place in the nineteenth century. The symmetry operators include rotation axes, glide and mirror

planes, and left- and right-handed screw axes which will simultaneously rotate and translate a three-dimensional object to create its clone in a different spatial position and orientation. The operators minimize the detail required to specify the spatial arrangements of patterns and objects that fill two-dimensional and three-dimensional space.

Symmetry is easily detected in common objects like a car wheel, where one nut and either a five- or sixfold rotation axis will generate other nuts holding the wheel in place. Wallpaper and quilts are useful for determining the minimal unit of pattern that must be specified within a cell, the combination of symmetry operations required to generate the complete pattern within a cell, and different shapes to outline two-dimensional cells that can be moved to fill the plane by translation. Quilts have well-defined symmetry operations and cells and identical shapes of nonidentical clones, if made from different fabrics. These are analogous to the so-called color space groups of crystallography, space groups that greatly increase the number of distinguishably different symmetries beyond the classical 230 by adding a fourth coordinate to the three space coordinates. This is done to encode a real difference that will be manifested in some property. The different directions of the magnetic moments of chemically identical atoms of an element such as iron provide an example of the need for representing a difference on the atomic level between cells that are otherwise identical.

Properties and models. Many physical and chemical properties were known, and models that accounted for them were proposed, before there was a means of getting inside crystals and mapping the components and mutual arrangements of components. Physical properties include the opacity and ductility of metals; the flakiness of mica; the hardness of diamond; the lubricity of graphite; double refraction from calcite, which gives two images of an object; pyroelectricity, where a thermal gradient builds up opposite charges on different faces; the piezoelectric effect in quartz; and the inherent anisotropy of crystals and the handedness of quartz, a handedness seen as either a counterclockwise or clockwise rotation of the plane of polarization of light. Polymorphs were known, as were chemical isomers, chemical compositions with the same molecular formula but different properties, such as ammonium cyanate (NH_4CNO) and urea [$CO(NH_2)_2$]. *See* BIREFRINGENCE; CRYSTAL OPTICS; DIAMOND; GRAPHITE; METAL, MECHANICAL PROPERTIES OF; PIEZOELECTRICITY; POLARIZED LIGHT; PYROELECTRICITY; QUARTZ.

Chemically identical but conformationally different (chair and boat) geometries had been proposed for cyclohexane. Steric hindrance was a concept, and structure models had been proposed and built to explain observed properties. Several were later confirmed by x-ray crystallography. Models based on the packing of ion spheres for sodium chloride (NaCl) and potassium chloride (KCl) were particularly insightful. Compositions, densities, and refractive indices in different directions had been measured, molecular formulas conjectured, and magnetic effects explored. A mathematically rigorous framework for mapping the internal mutual arrangements existed before there was a probe to enter into materials with light (radiation) that would send comprehensible signals revealing the species that built the structures, how these species were arranged, and what the arrangements implied regarding the forces between atoms. *See* CONFORMATIONAL ANALYSIS; STEREOCHEMISTRY.

Modern Crystallography

A suitable light or radiation to serve as a probe came with W. C. Roentgen's discovery of x-rays in 1895. The nature of x-radiation, whether waves or particles, was not immediately known, but the wavelength expected if x-rays were waves could be calculated to be of the order of 0.1 nanometer, comparable to the dimensions of atoms and 10^3–10^4 shorter than visible light. *See* ELECTROMAGNETIC RADIATION; X-RAYS.

Laue theory. The great insight of M. von Laue followed, leading to quantum physics and the reconciliation of the dual wave–particle nature of electromagnetic radiation, and also providing the means to get within the structures that chemists and mineralogists were modeling. Laue asked what would happen if x-rays really were waves; a crystal was really a regular, periodic, three-dimensional array of atoms; and the two entities were to meet. He concluded that the atoms would be excited to vibrate cooperatively, generating interference phenomena analogous to that of an optical diffraction grating. The same phenomenon can be observed by looking at a distant light through the mesh of a fine fabric or sieve. Spots are seen, which get larger as the mesh spacing gets smaller, regularly arranged about the direction of the incident beam.

Equations (1)–(3) must be satisfied to give

$$a(\alpha - \alpha_0) = h \tag{1}$$

$$b(\beta - \beta_0) = k \tag{2}$$

$$c(\gamma - \gamma_0) = l \tag{3}$$

constructive interference from a regular three-dimensional array of scattering lattice points; here a, b, c are the repeat distances of the space lattice; α and α_0, β and β_0, and γ and γ_0 are the direction cosines of the diffracted and incident beams; and h, k, and l are integers defining the order of a particular diffracted beam. Experiment showed all speculations to be correct: scattered x-rays combine with each other as waves do, crystals are three-dimensional arrays of regularly spaced scatterers, and the atoms and molecules of chemical models are real scatterers with 0.1-nm separation between their centers. Broadened diffraction spots reveal considerable distortion from lattice ideality, while very sharp spots come from crystals that are perfect within the resolution of the beam. *See* X-RAY CRYSTALLOGRAPHY; X-RAY DIFFRACTION.

Bragg spectrometer and modeling. The Bragg x-ray spectrometer uses the goniometer of Fig. 2 to study the x-ray spectrum, with the light replaced by x-rays, the collimator by a slit system, and the telescope by an ionization chamber. With mica and other crystals, the beam from a platinum target was shown to consist of continuous radiation, analogous to white light, and superimposed monochromatic radiation of various wavelengths that results from specific transitions between electron shells. The simple three-dimensional cubic chessboard model for structures of alkali halides (sodium chloride and potassium chloride), proposed many years before by optical crystallographers, was confirmed by W. L. Bragg by modeling all the spots found in Laue patterns, recorded on film by a stationary crystal, as well as the intensity differences seen when potassium chloride is substituted for sodium chloride. The following assumptions were used: (1) the mathematical planes of the lattice can act as x-ray mirrors; (2) the x-ray beam is heterogeneous; and (3) the scattering power of an atom is dependent upon its atomic weight.

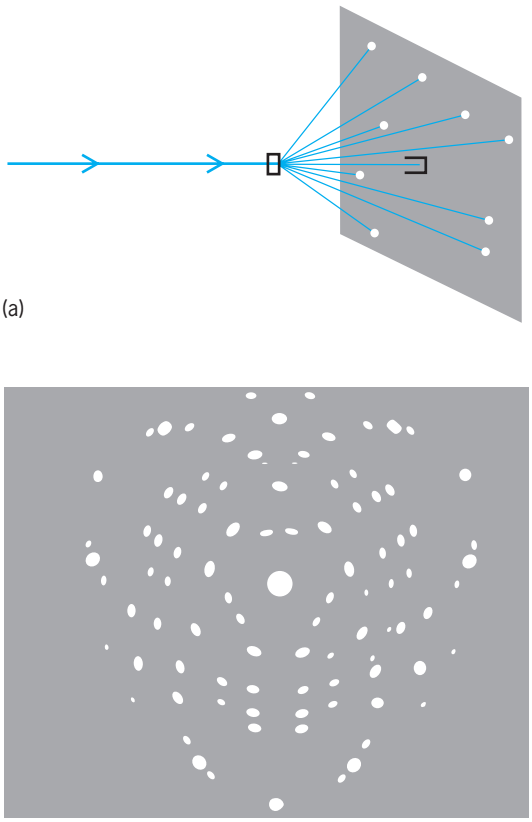
With a two-dimensional chessboard lattice arrangement, Avogadro's number, and the experimental densities, the distance between the sodium and chlorine atoms in a base-centered square cell containing two atoms (sodium and chlorine) per cell can be calculated. This distance is the sum of the radii of hypothetical atoms of sodium and chlorine. The required probe thus exists for entering structures and obtaining interpretable signals from atoms. The intensities of these signals can be understood in terms of interference from atoms whose distribution in space is modeled by idealized lattice and symmetry theory. The apparent sizes of scatterers—atoms, ions, or molecules—can be measured by assuming spheres in contact. The physical density and molecular formula determine the cell contents. Systematic work recording patterns for the same element in different chemical combinations and decoding the patterns by matching them with hypothetical models of scatterers, whose mutual arrangements are assigned from existing lattice and symmetry theory, gives values for the radii of spherical atoms.

Figure 7 shows the Laue pattern of calcium fluoride (CaF_2), recorded on flat photographic film. The spots are indexed by the projection method of Fig. 3. The distribution of spots reveals the geometric regularity and lattice symmetry of arrangement. The pattern has a threefold rotation axis (spots repeat every $360/3 = 120^\circ$) and a vertical mirror plane, normal to the film, which in effect reflects the spots. Spots on the left are seen on the right as if reflected by a mirror.

Bragg equation. The Bragg equation simplifies the Laue conditions to be met for constructive interference by using what had already been surmised for minerals: that crystals grow as if parallel planes were being added on.Equation (4) makes use of this to

$$\lambda = 2d_1 \sin\theta \tag{4}$$

decode observed patterns as if they originated from



(a)

(b)

Fig. 7. Laue pattern of calcium fluoride (CaF_2), which has cubic symmetry. The x-ray beam is along a body diagonal (111). (*a*) Experimental setup for generating the pattern. (*b*) Pattern on photographic film. (*Photographic Library, Palais de la Découverte*)

scattering by imaginary planes reflecting the x-ray light as if the planes were mirrors within a unit cell. Here, λ is the x-ray wavelength, $d' = d/n$, where d is the true spacing of lattice planes and n is the number of imaginary planes between them, and θ is the reflection angle.

For orthogonal crystals, Eq. (5) relates the Miller

$$\frac{1}{d^2} = \frac{h^2}{a^2} + \frac{k^2}{b^2} + \frac{l^2}{c^2} \tag{5}$$

indices of a plane, h, k, and l, the spacing d, and the lattice unit cell dimensions, a, b, and c. For the cubic system this simplifies to Eq. (6). More complex equations, involving angles, apply to lower symmetry.

$$\frac{1}{d^2} = \frac{h^2 + k^2 + l^2}{a^2} \tag{6}$$

X-ray spectra and synchrotron radiation. X-ray spectra have two origins. The beam ordinarily used in x-ray crystallography consists of sharp line spectra emitted by a target material when target atoms are ionized by removing electrons from innermost orbits. A limited heterogeneous spectrum of continuous (white) radiation is emitted during the sudden reduction of energy of the bombarding electrons. This is the radiation used in x-ray spectroscopy. The same phenomenon, reduction of electron beam

energy, but with enormous energies and highly parallel beams, is the source of synchrotron radiation. Synchrotron beams add an extremely powerful probe to x-ray crystallography. They are tunable x-ray sources; therefore, some atoms in a structure can often, in effect, be switched off by using wavelengths that will not excite them. In addition, the high energies give diffraction patterns in very short times. Therefore, the problem of degradation of biologically active materials during long x-ray exposures is eliminated. *See* BREMSSTRAHLUNG; SYNCHROTRON RADIATION; X-RAY SPECTROMETRY.

Technological and Scientific Applications

The sharp x-ray line spectra, characteristic of the bombarded element, are the primary probes for determining interior structural detail.

Techniques. Following the introduction of beams of this radiation, experimental and analytical techniques developed rapidly. Cameras with cylindrical film and enclosed powdered samples record all diffraction lines as arcs of concentric circles. This fundamental powder method has endured since 1917 and is now employed with improved beam purity and optics, improved diffractometers which couple sample and detector rotation, electronic detection, rapid sequential recording, and computer indexing programs that provide patterns of compounds, mixtures of phases, and dynamic changes that occur when crystals are subjected and respond to external stress. The method is applied to single crystals, polycrystalline aggregates, and multiple-phase mixtures, randomly disposed either in space or in geometrically designed composite materials. *See* X-RAY POWDER METHODS.

Data collection. The x-ray patterns of virtually every material that can be handled have been recorded and indexed. Comprehensive collections of experimental and model calculation results for metals, minerals, and inorganic compounds are kept current and accessible through the Joint Committee for Powder Diffraction Study (JCPDS). More sharply focused data files for structures of organic and biological interest are also extensive and report molecular configurations.

Basic structures and properties. Counting the number of neighbors about an atom and measuring the distances between their centers from the diffraction patterns of elements and compounds yield the basic elements of structure: atoms, molecules, ions, and the modern concept of the chemical bond. These procedures also yield explanations for some of the properties of materials.

Many metals fill space with a close-packed periodic arrangement of spheres in contact, but the space is really quite empty. The positively charged atom cores repel each other electrostatically and maintain large equilibrium separations through a sea of mobile electrons, which is the source of metallic conductivity. The mobility of the electron sea, the emptiness of the structures, and the mutual repulsions of the atom cores explain metallic ductility. The effective radii of atoms of the elementary metals are seen to be much greater than those found when the metals are in combination with other elements. Effective radii are also different in compounds having different numbers of neighbors around the same element, smaller for fewer near neighbors, larger for more. L. Pauling's theory of the chemical bond is based upon such crystallographic data. *See* CHEMICAL BONDING; ELECTRICAL CONDUCTIVITY OF METALS; FREE-ELECTRON THEORY OF METALS.

Relation of structure and composition. Decoded patterns of materials of very different compositions and even from widely different origins show surprising similarities when viewed from within the framework of their structural architectures. For example, crystallographic patterns reveal much similarity of structural architecture between fibrous protein materials, such as wool, hair, and the myosin of muscle, which have very different compositions. *See* FIBROUS PROTEIN.

An example of materials of different origins is provided by silica (SiO_2) and ice (H_2O), which are extraordinarily alike in their polymorphic forms, the geometry of their building units, and the connectivity schemes that link the units to form the solids. There are four tetrahedrally arranged oxygen atoms about a silicon atom in silica and four hydrogen atoms about an oxygen atom in all the forms of ice. The variety of possible forms is seen as the many spatial arrangements that can be made by connecting each tetrahedron at its corners to other tetrahedra. *See* SILICA MINERALS; WATER.

In contrast to these architectural similarities between chemically different substances, other patterns show different structural architectures for identical compositions. Furthermore, fundamental differences in apparent size (as mentioned above) and mutual arrangements are observed among atoms of the same element. These differences depend upon the composition that the atoms are part of and also on the conditions of formation. The division between atomic, ionic, and covalently bonded arrangements is established by systematic measurements on the periodic lattice that models the observed diffraction patterns.

Studies of response to change. Modeling crystallographic results with space-filling unit cells of the proper density, independent of the mutual arrangement of atoms or molecules within the cell, provides the means for following, in the laboratory, the degrees of freedom of the overall structure, which are the types of adjustment that the structure makes to accommodate change without disruption. These are determined in several ways. Static determinations, which involve varying the composition systematically and comparing the patterns and densities of a family of compositions, reveal solid solutions and phase relations important to materials development. Dynamic determinations involve changing the temperature or pressure between a sequence of pattern recordings. Thermal expansion and compressibility are important both in developing new materials and properties and in understanding geological processes.

Dynamic studies. The foundation of space-filling lattice theory allows internal changes to be studied under laboratory conditions. Even when the details of mutual rearrangements are not known exactly, the size and shape of unit cells with the proper physical density can often be obtained from indexed patterns, density measurements, and a correct chemical formula. Subjecting a polycrystalline powder to external sources of stress is a dynamic way of focusing on structural details, such as the directions of strongest bonding, that are strongly implied but not directly measured. Heating, cooling, and the application of pressure cause changes in the size and shape of the unit cell as responses to the stress applied.

Changes in relative diffracted intensities indicate that the phase relationships that build up the total intensity are changing. This is often a signal that new modes of interatomic motion have been generated. Dynamic applied crystallography reveals many fundamental phenomena. A solution of elements that were initially uniformly distributed can coalesce as one-dimensional strings or two-dimensional platelets in specific crystallographic directions within the host lattice. A lively pattern of cooperative movements at the atomic level of structure can be observed crystallographically.

Crystals always sense and respond to environmentally imposed stress. They strive to equilibrate their structures to the new environment. Permanent and irreversible processes, as well as reversible ones, occur frequently. Detecting these is critical to the fundamental understanding of the solid state as well as the development of compositions and processes that produce materials that survive under conditions and environments of use.

Structure of glass. The surprising observation that there are true glasses that fill space uniformly with nonperiodic arrangements stimulated a reexamination of scattering theory and improvements in the detection of weak signals. The atomic structure of glass is explained by radial density distributions of scatterers in physical space such that all identical atoms in the structure have the same number and relative arrangement of their nearest neighbors. A geometrically irregular network that distributes spherical concentric shells of appropriate scatterers at larger distances accounts for the background pattern of diffuse scattering. A physical model of identical tetrahedral units, completely connected at their vertices and randomly arranged in space, fits the experimental scattering data. *See* AMORPHOUS SOLID; GLASS.

Living systems and other challenges. The dynamics of living systems, the difficulties in distinguishing light elements, and the inherent ambiguities of measuring, decoding, and mapping are continuing challenges. Major achievements of crystallography include the determination of the structures of deoxyribonucleic acid (DNA), proteins, other biological compounds, and boranes; the development of direct methods of phase determination; and the determination of the structure and mechanism of a photosynthetic center. *See* BORANE; DEOXYRIBONUCLEIC ACID (DNA); PHOTOSYNTHESIS; PROTEIN.

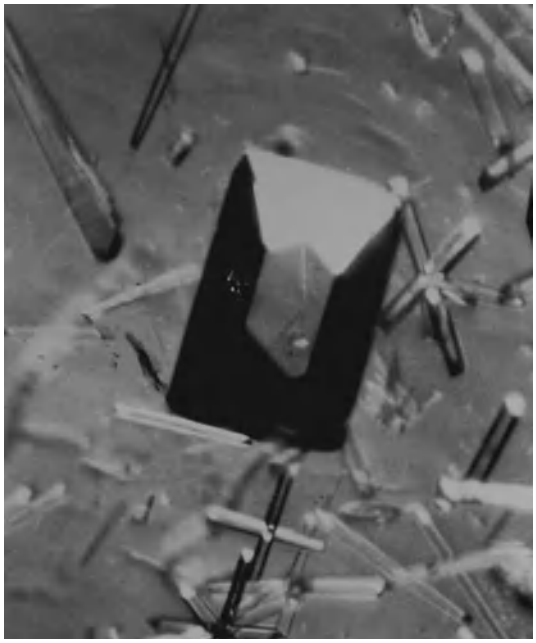


Fig. 8. Large porcine elastomer crystal grown in Spacelab (1988). (*Charles Bugg*, *University of Alabama*)

The problems encountered in the study of biological systems include the large atomic populations of these systems and the small scale of resolution needed to observe how components arrange themselves and how their configurations readjust, in response to subtle signals, with great speed and over extensive distances. **Figure 8** illustrates an ongoing challenge in the study of living systems: producing good crystals from material that does not ordinarily crystallize, and could not perform its biological function if it did crystallize. Although the crystal of porcine elastomer is considered large, it is 0.5 mm in the largest dimension, about 1/100 the size of the crystals in Fig. 1. It grew aboard Spacelab in 1988. Since convection forces on Earth disrupt the growth process, a gravity-free environment is required to grow a crystal of this protein, see inside it, and map its internal architecture. Only then can the rapid communication and readjustments that take place in living systems be modeled and understood. *See* SPACE PROCESSING.

Advanced probes and detailed modeling. Many kinds of radiation are used to enter within and excite signals from a crystal structure. All are employed to determine detail and fix positions for molecules and atoms within a space that is modeled as the real interior. Success in modeling depends upon extracting regularities and symmetries from signals that originate in the inner space of a material. These regularities are used to reconstruct both the summation of many amplitudes that yields the observed intensities, and the kind and mutual arrangement of the scatterers that locate these intensities in their observed positions. Relating these arrangements to internal activities is the goal of all crystallography. *See* ELECTRON DIFFRACTION; NEUTRON DIFFRACTION.

Icosahedral crystals. Conditions of formation in some three-component metal melts produce very small and morphologically well-defined icosahedral crystals. The idealized interior space of these crystals may not be correctly mapped in terms of diffraction patterns referenced to the three vectors of a Bravais lattice, but require so-called higher-dimensional spaces. *See* QUASICRYSTAL. Doris Evans

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Ctenophora

A phylum of exclusively marine organisms, formerly included in the jellyfish and polyps as coelenterates. These animals, the so-called comb jellies, possess a biradial symmetry of organization. They are characterized by having eight rows of comblike plates (*ctenos* = comb) as the main locomotory structures. Most of these animals are pelagic, but a few genera are creeping. Many are transparent and colorless; others are faintly to brightly colored. Many of these organisms are hermaphroditic. Development is biradially symmetrical, with a cydippid larval stage. Five orders constitute this phylum as follows:

Phylum Ctenophora
Order: Cydippida
Lobata
Cestida
Platyctenida
Beroida

Morphology. The body is gelatinous and extremely fragile; its form may be globular, pyriform, or bell- or helmet-shaped. Some species resemble a ribbon. Their size ranges from $^1/_8$ to 20 in. (3 mm to 50 cm). All are biradially symmetrical. The axis of symmetry is determined by the mouth and the organ of equilibrium, or statocyst (**Fig. 1**). The mouth leads into the flattened, elongated pharynx. The sagittal plane is thus referred to as the stomodeal plane. The other plane of symmetry is perpendicular to the sagittal plane and is marked (Beroida excepted) by tentacles. This is known as the tentacular, transverse, or equatorial plane. Eight meridional comb-plate rows or ribs stretch from the aboral pole on the surface of the body.

Comb plates. These are the most characteristic structure of the ctenophores, possessed by all members of the phylum. Each plate consists of a great number of very long related cilia. Successive plates are arranged as meridional ribs or costae which extend from the aboral pole toward the mouth. Each rib is connected to the statocyst by a thin ciliary furrow. Of the eight ribs, four are stomodeal and are located on both sides of the stomodeal plane. The remaining four are subtentacular ribs and are situated on both sides of the tentacular plane. Comb rows are secondarily modified in some specialized ctenophores, and ribs may be entirely absent in the adult stages of a few genera of Platyctenida.

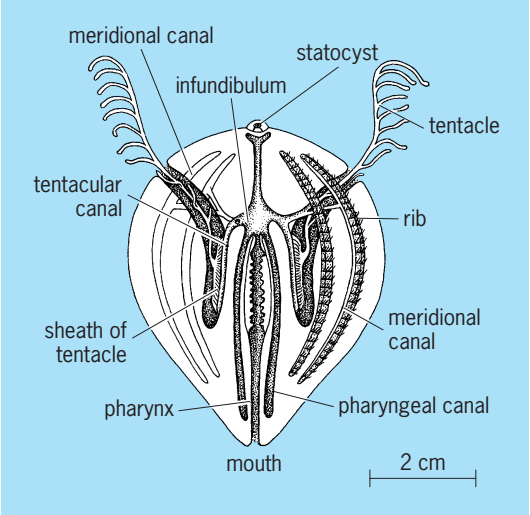

Fig. 1. Structure of a cydippid ctenophore.

Tentacles. With the exception of Beroida, all ctenophores possess two highly extensile tentacles bearing many lateral branches (tentilla). The base of each tentacle is anchored on the bottom of a sheath which is formed by a deep depression in the body surface, and into which the tentacular apparatus may be completely retracted. Tentacles and tentilla are thickly covered with adhesive cells (colloblasts) which are exclusive to the phylum. Each colloblast is connected by the nervous system. A colloblast consists of a cell which looks like a mushroom (**Fig. 2**), the foot of which is firmly anchored in the fibrillar jelly surrounding the neuromuscular axis of the tentillum. This foot contains a helicoidal twisted thread which reaches the expanded head of the colloblast. Here, the helical thread divides into a great number of radii. Each radius terminates by a granule of glue. When a prey collides with a tentillum, some colloblasts burst and their granules stick the prey to the helical threads. These threads apparently act as shock absorbers of the captured prey's movements rather than as springs projecting the head. Tentacles are organs in continuous growth. Their formative tissue is located at their base, in the bottom of the tentacular sheath. In the adult stages of the Lobata and Cestida, the main tentacles, as described above, are reduced, and a series of tentilla with colloblasts develop on the margin of the mouth.

Gastrovascular system. The pharynx is connected aborally with a small chamber, the stomach, or infundibulum, from which a canal system common to all ctenophores ramifies throughout the mesoglea. The most important of the canals are the meridional

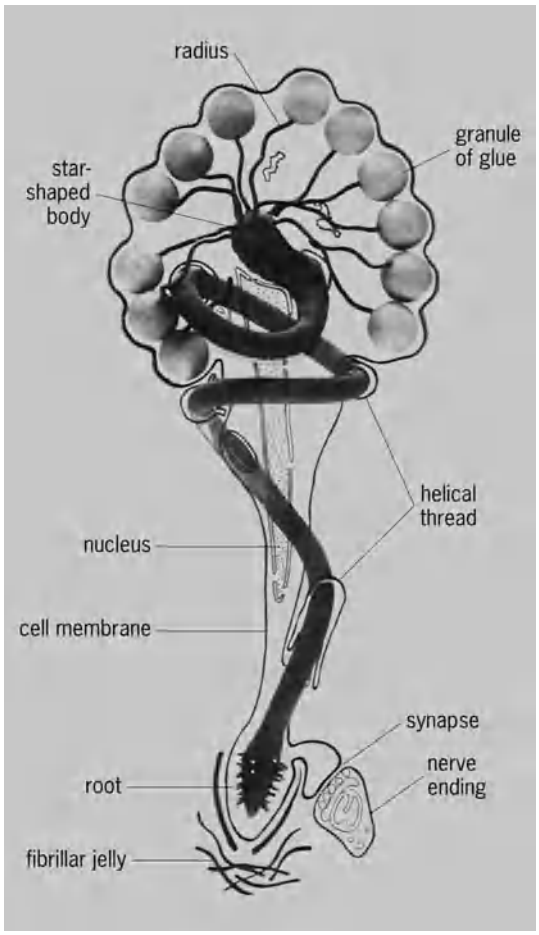

Fig. 2. Colloblast from the tentacle of a ctenophore.

canals which underlie the ribs, the pharyngeal (stomodeal) canals which run along the pharynx, and the tentacular canals (when tentacles exist) which go to the base of the tentacles. While the pharynx has an ectodermal lining, the stomach and the canal system are lined with endodermal epithelium which includes digestive cells. The peripheral canals bear special structures, the cell rosettes, consisting of the superposition of two circles of eight ciliated cells. The peculiarity of these structures lies in a contractile diaphragm which allows complete opening or closing between the mesoglea and the lumen of the gastrovascular canals. The cell rosettes may be considered peculiar excretory or hydrostatic organs, since reversible liquid flows were observed through them with changes in seawater salinity.

Mesoglea. The maintenance of shape in the ctenophores depends on the mesoglea. This inner jelly, very rich in water (95%) is mainly made of protein-carbohydrate complexes. Collagen was biochemically detected, but has not been related to definite mesogleal structures. The mesoglea is crossed by numerous smooth muscle cells which are among the longest (2 in. or 5 cm) of the animal kingdom. Nerve processes connecting the muscle cells have been observed in the mesoglea, where scattered mesenchymal cells are also found.

Nervous system and sensory organs. Ctenophores possess a nervous system organized in a subectodermic synaptic network, which is more condensed along the ribs, under the statocyst, and in the tentacles. There is some evidence for the existence of a cholinergic system together with a monoaminergic one. Although intramesogleal neurons exist, innervation of the endoderm has not been observed. Sensory cells sensitive to water movement are located over the entire surface of the body and are connected to the nervous network. They are most abundant along the tentacles, the lips, and at the aboral pole. The nerve net propagates a localized excitation over the entire body of the animal and complex behavior patterns may take place (feeding behavior, for example). The statocyst is situated in a hollow of the aboral pole. The main feature of this equilibrium organ (**Fig. 3**) is the statolith, a mass of cells containing calcareous granules, held by and linked to four S-shaped tufts of cilia, the balancers. Each balancer, interradially located, is connected with the subpharyngeal and subtentacular costae of the same quadrant by the two corresponding ciliary furrows. The floor of the hollow contains secretory cells involved in the formation of the statolith and sensory cells, presumably photoreceptors and baroreceptors. A fence of incurved cilia forms a transparent dome over this aboral sensory area. Paired structures called the polar fields lie close to the statocyst, in the pharyngeal plane, and consist of linear or crescentic ciliated areas, the function of which is unknown.

Gonads. Many ctenophores are hermaphroditic. Gonads of endodermal origin develop in the outer side of the wall of the meridional canals or their branches. Both the ovaries and spermaries generally develop as continuous tracts of germ cells. The ovaries occupy the perradial and the spermaries the interradial side of the wall. Contrary to the spermatozoa of Porifera and Cnidaria, the ctenophore spermatozoon possesses a well-differentiated acrosome. Gonducts seem to be of general occurrence. Functional gonads in the larval stage occur throughout the phylum.

Embryology. Fertilization usually occurs in seawater. Pelagic ctenophores are self-fertile, but cross-fertilization might also take place inside a swarm of ctenophores. Ctenophore eggs provide the unique example of a biradial pattern of cleavage (**Fig. 4**). The blastomeres are arranged in respect to the future

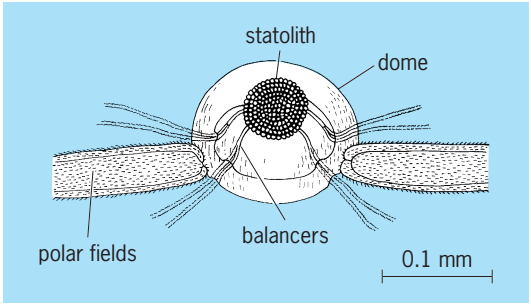

Fig. 3. Features of a statocyst.

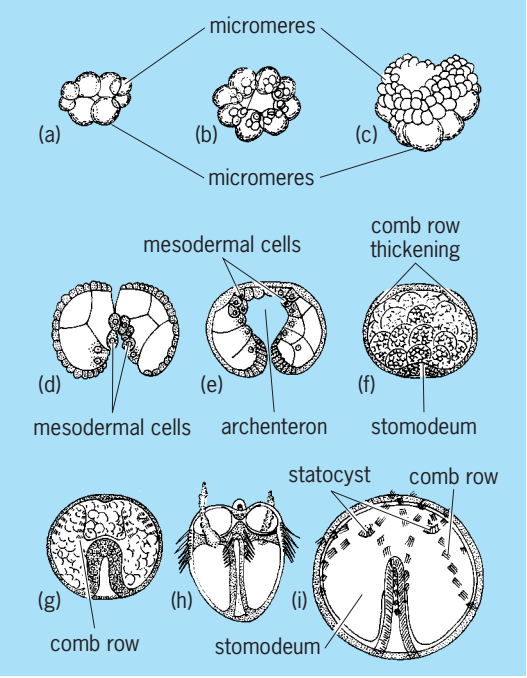

Fig. 4. Embryology. (*a*) Eight-cell stage. (*b*) Micromeres, viewed from above. (*c*) Later stage in division of micromeres. (*d*) Gastrula, section. (*e*) Invagination. (*f*) Stomodeum and thickenings for comb rows. (*g*) Stomodeum invaginated; comb rows differentiated. (*h*) Cydippid larva. (*i*) Embryo with statocysts and comb rows.

two main planes of symmetry of the animal. Gastrulation is first epibolic (the macromeres are gradually covered by the multiplication of the micromeres), but ends with invagination of the last micromeres produced by macromeres. Development of the embryo is determinate. The claimed endodermal formation of a true mesoderm, giving rise to muscular and mesenchymal cells, is still debatable. The earliest larva formed in the development of all types of ctenophores is the cydippid, with a globular body, a pair of primary tentacles (or rudiments), and eight costae of comb plates. In the Lobata, Cestida, and Platyctenida, the larva undergoes secondary changes in later development.

Physiology. The ctenophores move through the sea with the oral pole forward, using comb plates as oars. The beating of successive plates occurs along each costa by metachronal waves and is initiated for the two costae of the same quadrant by the corresponding balancer of the statocyst. Balancer and comb-row activities and some muscular contractions are integrated by the nervous system.

Almost all ctenophores are luminescent. This capacity, linked to a calcium-dependent photoprotein, is assumed by specialized cells underlying the comb plates.

The ctenophores feed on zooplankton. Prey are trapped with the tentacles (Cydippida, Cestida) or oral extensions of the body (Lobata), or are engulfed directly through the widely opened mouth (Beroida). Extracellular digestion begins in the pharynx, where gland cells secrete enzymes. The partially digested material is then carried throughout the gastrovascular system, where it is phagocytized by digestive cells.

Ctenophores have high powers of regeneration. Asexual reproduction (in a few platyctenid genera) is by regeneration of an entire organism from a small piece of the adult body.

Ecology. Ctenophores are themselves important plankton organisms. They are quite common and have a worldwide distribution in all seas, where they can appear in enormous swarms. Some genera stand great changes in the seawater salinity. Because of their voracity as predators of zooplankton, they play an important role in plankton equilibrium and in fisheries.

Phylogeny. The Ctenophora constitute a well-defined phylum. Among its orders, the Cydippida are undoubtedly the most primitive, with radiating evolution of the other orders. The ancestor of the phylum is possibly to be found among the Trachylina (Cnidaria, Hydrozoa). However, because of such forms as *Coeloplana* (Platyctenida), ctenophores have also been suspected of affinity with the polyclad Turbellaria. *See* POLYCLADIDA; TRACHYLINA.

J.-M. Franc

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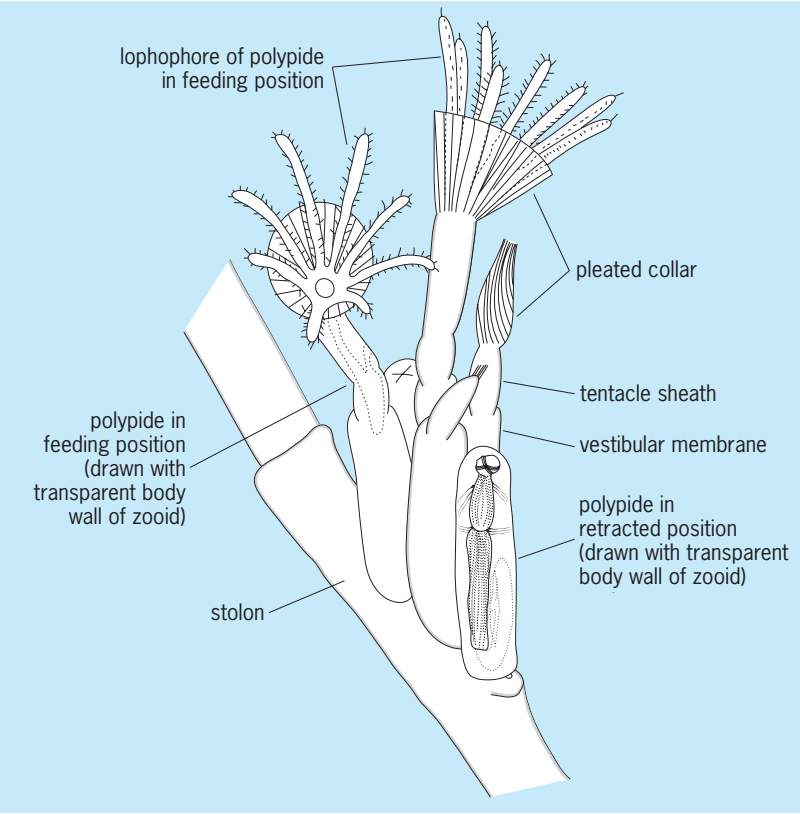
Ctenostomata

A paraphyletic order of Bryozoa in the class Gymnolaemata. Ctenostomes have entirely cuticular body walls, and colonies consist of feeding zooids (individual bryozoans) only or of feeding zooids plus spine- or stolon-shaped zooids. *See* BRYOZOA; GYMNOLAEMATA.

Morphology. Feeding zooids possess tentacles that collectively form a circular, feeding, bell-shaped lophophore centered on the mouth. The mouth is located atop a cuticular tube enclosing the initial portion of the U-shaped digestive tract, and the terminal portion of the digestive tract ends as an anus part of the way up the cuticular tube. When the lophophore is retracted into its protected position within the main body of the zooid, the cuticular tube inverts and surrounds the lophophore as a tentacle sheath, and the opening through which it protrudes to feed is closed by a sphincter muscle. At the muscular diaphragm of the inverted cuticular tube, ctenostomes have a cuticular collar referred to as either a setigerous or pleated collar (see **illustration**). The collar extends as a cone around the lower parts of the lophophore when the zooid is feeding but



Cluster of zooids of the ctenostome genus *Amathia*.

at other times is collapsed into a narrow, longitudinally pleated tube that extends from the region of the sphincter muscle at the outer tip of the zooid (see illustration). Many, perhaps all, cuticular collars have several longitudinal rods of thickened cuticle that project from the outer end of the cone to give a comblike appearance; hence the name of the group (cteno- = comb; stome = mouth).

Most ctenostomes grow as encrusting or erect, ramifying, thin branches. Some ctenostomes consist entirely of feeding zooids (for example, *Victorella*), others of individual feeding zooids (for example, *Bowerbankia*) or feeding zooid groups (for example, *Amathia*; see illustration) connected by zooids specialized as stolons. Other ctenostome colonies, such as *Alcyonidium*, consist entirely of densely packed feeding zooids that form flexible encrustations or thick, erect, branched growths. A few branching ctenostomes are endolithic borers, creating their living space by dissolving carbonate substrates such as shells just below the surface.

Life cycle. Colonies are hermaphroditic. Fertilized eggs are usually brooded as embryos in the space normally occupied by the retracted tentacles of feeding zooids. In other species, brooding may be within specialized invaginations or within the coelom. A few species release fertilized eggs directly into the sea. The free-swimming larvae attach to a substrate for metamorphosis (including absorption and histolysis of all larval organs), resulting in the formation of the first feeding zooid of the colony, from which all subsequent zooids within the developing colony are budded.

Most ctenostomes are marine, although several species are tolerant of lower salinity than any other marine bryozoans and a few live in freshwater. Ctenostomes range from the low intertidal zone to the abyss.

History and classification. The fossil record of endolithic ctenostomes extends back to the Late Ordovician. The lack of a mineralized skeleton diminishes chances of preservation of nonboring ctenostomes. However, where rapidly growing skeletons of organisms such as oysters extend over ctenostomes, the ctenostome shapes may be molded in surprising detail on the undersurfaces of the overgrowing organism by a process called bioimmuration. Bioimmured ctenostomes are known back to the Jurassic. *See* JURASSIC; ORDOVICIAN.

Ctenostome bryozoans are relatively primitive in their lack of calcification, small number of zooid types, lack of brooding in some species, and other features. The oldest known Cheilostomata, of the Late Jurassic, have zooidal shapes and encrusting colony growth habits that are close to morphologies known in extant ctenostomes and apparently are derived from ctenostomes. The class Stenolaemata may have originated from ctenostomes by splitting off the muscle-bearing mesodermal layer from the original outer body wall as calcification rigidified the lateral body walls of cylindrical zooids, leaving the flexible terminal wall and the mesodermal layer of the lateral walls involved in protrusion of tentacles to the feeding position. *See* STENOLAEMATA. Frank K. McKinney

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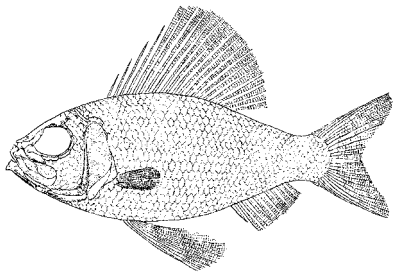
Ctenothrissiformes

A small order of teleostean fishes that are of particular interest because they seem to be on or near the main evolutionary line leading from the generalized, soft-rayed salmoniforms to the spiny-rayed beryciforms and perciforms, dominant groups among higher bony fishes.

Ctenothrissiformes lack fin spines but they share the following characters with primitive beryciforms: pelvic fins that are thoracic, with seven or eight rays, and pectoral fins that are placed well up on the



Cretaceous ctenothrissid, *Ctenothrissus radians*; length 10–12 in. (25–30 cm). (*After C. Patterson, Phil. Trans. Roy Soc. London, Ser. B, Biol. Sci., no. 739, vol. 247, 1964*)

flank (see **illus.**); a subocular shelf; a scaly opercle; a mesocoracoid lacking; a similar premaxilla; maxillary teeth; an orbitosphenoid; and 17 branched caudal rays. Except for the last three these features persist in generalized Perciformes and indicate a major evolutionary trend away from typical soft-rayed fishes. The scales are ctenoid or cycloid, and there are nine branchiostegals. *See* SCALE (ZOOLOGY).

The few fishes confidently assigned to the Ctenothrissiformes, *Ctenothrissus* (Ctenothrissidae) and *Aulolepis* and *Pateroperca* (Aulolepidae), are known as Upper Cretaceous marine fossils from Europe and southwestern Asia. *See* BERYCIFORMES; PERCIFORMES; SALMONIFORMES. Reeve M. Bailey

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Cube

A parallelepiped whose six faces are all squares. The cube (see **illus.**) is one of the five regular solids known to the ancient Greeks, who proposed the

A cube, a type of parallelepiped.

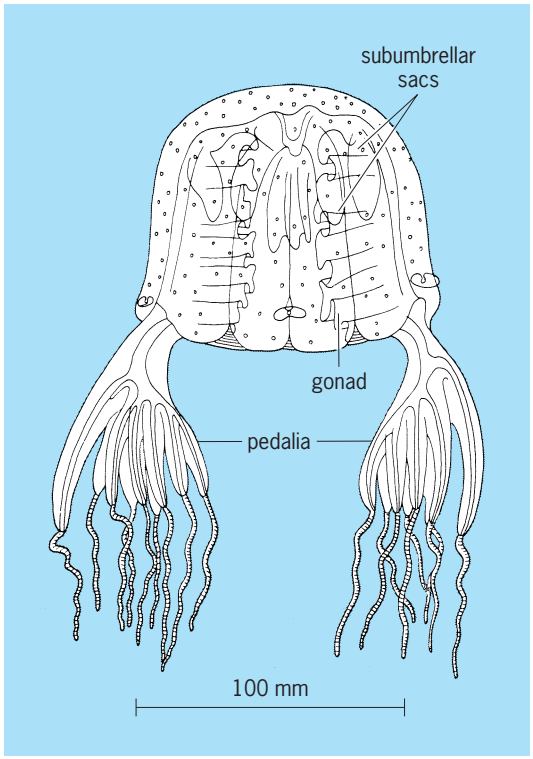
famous problem (now proved to be unsolvable) of constructing, with the use of the compass and unmarked straightedge alone, a cube whose volume is twice that of a given cube. The edge of the desired cube may be found by use of conic sections. A cube has eight vertices and twelve edges. Each vertex is on three edges and three faces, each edge is on two vertices and two faces, and each plane is on four vertices and four edges. *See* PARABOLA; POLYHEDRON; REGULAR POLYTOPES. Leonard M. Blumenthal

Cubeb

The dried, nearly ripe fruit (berries) of a climbing vine, *Piper cubeba*, of the pepper family (Piperaceae). This species is native to eastern India and Indomalaysia. The crushed berries are smoked, and the inhaled smoke produces a soothing effect in certain respiratory ailments. Cubebs are used in medicine as a stimulant, expectorant, and diuretic. Medicinal properties of cubeb are due to the presence of a volatile oil which formerly was thought to stimulate healing of mucous membranes. *See* PIPERALES. P. D. Strausbaugh; Earl L. Core

Cubozoa

An order of the class Scyphozoa. They are good swimmers and their distribution is mostly tropical. The umbrella is cubic, and a pedalium, a gelatinous leaf-shaped body, is present at the base of each ridge of the exumbrella (see **illus.**). A well-developed



Cubomedusan, *Chiropsalmus quadrumanni*. (*After L. Hyman, The Invertebrates, vol. 1, McGraw-Hill, 1940*)

tentacle, or cluster of tentacles, arises from each pedalium. The sensory organs are on the perradii. They are the most highly developed in the Coelenterata, each one having six ocelli and a statocyst. The manubrium, a peduncle connecting the mouth with the stomach, has four lips. The cubic stomach connects with four radial pockets, and between these

pockets are the interradial septa. The leaf-shaped gonads develop from the septa and extend into the pockets. No connecting stage between polyp and medusa is known. *See* SCYPHOZOA. Tohru Uchida

Cuculiformes

An order of birds having zygodactyl feet (two toes are in front and two behind) in which the fourth toe is reversed, but not having any close relationship to other birds with this foot structure, such as the parrots or piciforms. The relationship of the touracos to the cuckoos has been strongly debated and is still unresolved. However, this conclusion is very controversial because the hoatzin possesses an anisodactyl pattern of the toes (three toes are in front and one is in back); thus, many authors still place the hoatzin in its own order next to the Cuculiformes. *See* OPISTHOCOMIFORMES; PICIFORMES; PSITTACIFORMES.

Classification. The order Cuculiformes is divided into two suborders: Musophagi, containing the single family Musophagidae (touracos; 18 species), and Cuculi, with the large and diverse family Cuculidae (cuckoos; 129 species). Cuculidae is divided into six distinct subfamilies, some of which have a limited distribution, such as the couas (Couinae; 10 species) which are confined to Madagascar.

Fossil record. The fossil record of the Cuculiformes is poor and comprises only scattered finds. Although extant members of the Musophagidae are found only in Africa, several fossils of this family have been found in the Oligocene and Miocene of Europe, indicating that this group once had a wider distribution.

Touracos. Touracos are African woodland birds with a short- to medium-length crest on their heads, and a loose green, gray, or brown plumage with patches of bright yellow, red, or violet. Most species feed largely on fruit, although several species also consume insects. One species, the Great Blue touraco (*Corythaeola cristata*), also eats leaves, especially late in the day; it feeds regurgitated leaves to its young after their first week of life. The green pigment (touracoverdin) in the feathers is unique and is the only true green pigment known in birds; the red pigment (touracin) is water-soluble and copper-based and is unique to birds. The pigment in the feathers is protected from washing away by overlying feathers. The nest is a bulky platform placed in a tree. Incubation and care of the chicks is shared by both parents. The young leave the nest at an early age (about 10 days) and crawl about the trees before being able to fly.

Roadrunner (*Geococcyx* sp.). (*Photo by Marguerite Gregory, © 2004 California Academy of Sciences*)

Cuckoos. Cuckoos are a morphologically and behaviorally diverse worldwide group, as indicated by the six recognized subfamilies, of which three are found only in the Old World and three only in the New World. Their size varies from small (30 g or 0.07 lb) to large (channel-billed cuckoo, *Scythrops novaehollandiae*, at 600 g or 1.32 lb). They are mostly arboreal, but members of three subfamilies are largely or entirely ground-dwelling; some, such as the American roadrunners (*Geococcyx*; see **illustration**), are fast runners, with top speeds of about 15 mi/h (25 km/h), and rarely fly. Most cuckoos are secretive birds, sulking in heavy vegetation. They are strong fliers but do not fly often. Temperate species are migratory. Cuckoos feed on insects, small vertebrates (including poisonous snakes taken by roadrunners, *Geococcyx*), and other animals, and less commonly on fruit; no species feeds entirely on fruit. Most are solitary, but the New World anis and guiras (Crotophaginae) live in flocks and build a communal nest housing several females.

All species of Old World cuckoos (Cuculinae) are nest parasites, as are three species of American ground-cuckoos (Neomorphinae). The common cuckoo of Europe is the most studied nest parasite, and has many adaptations in both the adult female and the young to ensure success in having the host species rear the young. Females of these species belong to different "egg varieties," in which the eggs resemble those of different hosts. The eggs laid by cuckoo females in host nests are small compared to the size of the adult bird. Incubation is short, and the young cuckoo pushes the other eggs or young out of the nest. The host parents feed the parasitic chick, which grows to a gigantic size compared with the adult host. Most American species build their own nests and care for their own young.

Cuckoos are valuable as insect eaters, especially of toxic and hairy caterpillars rejected by other birds. The European cuckoo has held an important place in European folklore, literature, and music, as has the roadrunner in the southwestern United States and in Mexico. *See* AVES. Walter J. Bock

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Cucumber

A warm-season annual cucurbit (*Cucumis sativus*) native to India and belonging to the plant order Campanulales. *See* CAMPANULALES.

Production. The cucumber is grown for its immature fleshy fruit which is used primarily for pickling and for slicing as a salad (**Fig. 1**). The development of

Fig. 1. Cucumber fruits. (*Asgrow Seed Co.*; *Upjohn Co.*)

gynoecious (bearing only female flowers) and hybrid varieties (cultivars) with increased disease resistance is resulting in a continuing change in the varieties of both pickling and slicing cucumbers. In addition, the advent of mechanical harvesting and chemical weed control for pickling cucumbers is altering traditional methods of culture, for example, higher plant populations per acre and single harvests instead of multiple pickings. Important cucumber-producing states are Florida and South Carolina for fresh market and Michigan and Wisconsin for pickling. *See* SQUASH.

H. John Carew

Diseases. Diseases of cucumbers reduce plant growth, fruit yield, and product quality in both field and greenhouse plantings. Because cucumbers are grown all months of the year, it is necessary to guard against seeding in unfavorable environments and to control seed decay and seedling blight.

Bacterial diseases. Bacterial wilt is a serious disease in the midwestern, north-central, and northeastern regions of the United States, but it is not important in the South or West. Plants infected with the causal bacterium *Erwinia tracheiphila* first show a wilting of one or a few leaves which remain green, but later all the leaves wilt and the plant dies. A white, stringy, viscid bacterial ooze shows in freshly cut wilting stems. Bacteria-infested 12-spotted and striped cucumber beetles infect the plant while feeding upon it. The bacteria multiply in the plant and become distributed throughout the vascular system. The bacteria overwinter in the body of the insect. Because the only way in which the plants can become infected is through the feeding of the cucumber beetles, control depends upon early and prompt destruction of these pests by insecticides. Antibiotics have sometimes proved useful against the bacterium.

Angular leaf spot, incited by the bacterium *Pseudomonas lachrymans*, causes water-soaked spots on leaves and bordered, irregular lesions on fruits. The lesions later develop into a brown firm rot extending into the flesh (**Fig. 2**). In moist conditions, bacteria ooze from the lesion in "tears" which dry into a whitish residue. The causal bacterium overwinters on infected vine refuse, is seed-borne, and is spread by rain and surface water. Control consists of using California seed, which is low in infection, soaking it for 5 min in a mercuric chloride solution (1 part in 1000 parts of water), rinsing the seed in water, and planting promptly. Fixed copper sprays can achieve control, but must be used carefully as they can be phytotoxic to young plants. There are some resistant varieties available.

Fungal diseases. Downy mildew is a destructive disease in the eastern and southern states where the weather is warm and moist, is less damaging in the north-central states, and rarely occurs in the Southwest. It is incited by the leaf-inhabiting fungus *Pseudoperonospora cubensis*, which causes angular, yellowish spots; the spots later turn brown when the

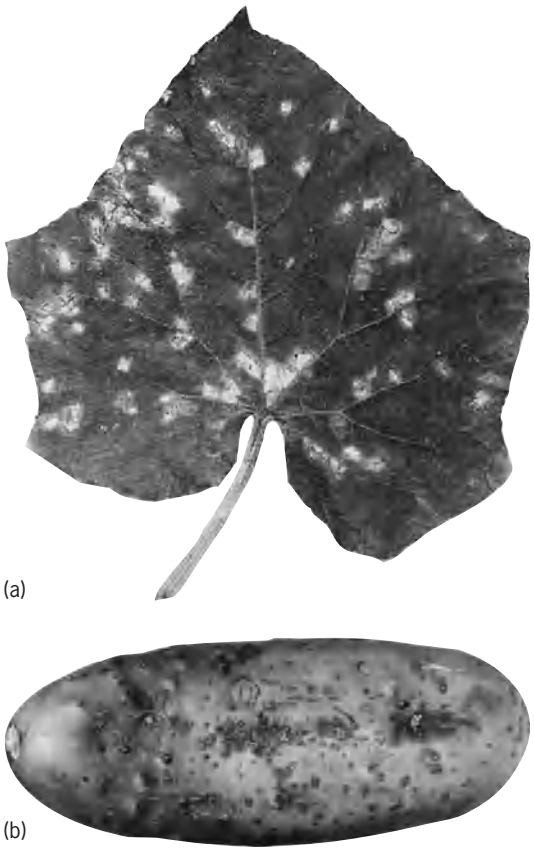


Fig. 2. Angular leaf spot disease of cucumber caused by bacterium *Pseudomonas lachrymans*. (*a*) Leaf spots. (*b*) Fruit lesions. (*Courtesy of J. C. Walker*)

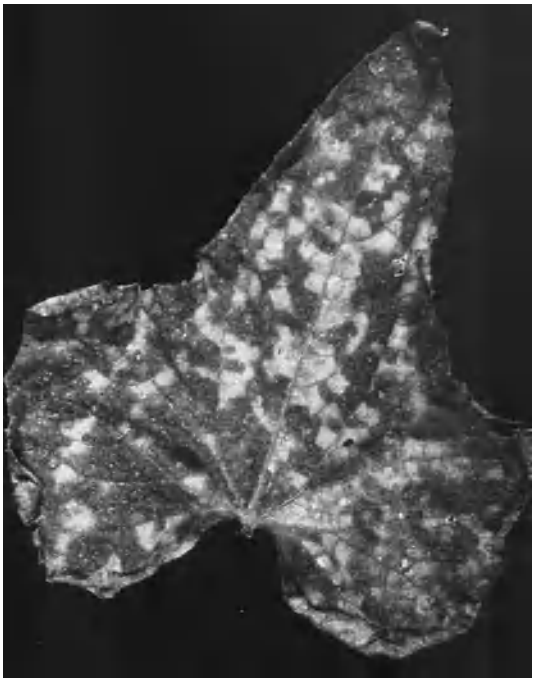


Fig. 3. Downy mildew of cucumber leaf showing yellow angular blotches induced by *Pseudoperonospora cubensis*. (Courtesy of D. E. Ellis)

leaves shrivel (**Fig. 3**). In moist conditions, the lower leaf surface is covered with the fruiting layer of the fungus. Entire vines may be killed. The loss of foliage interferes with normal flower set and fruit development. Maturing fruits fail to color properly, are tasteless, and are usually sunburned. Planting of resistant varieties, such as Ashley, Palmetto, Palomar, Santee, and Stono, and treating plants with copper fungicides will provide maximum protection from downy mildew.

Scab is a serious disease in the northern and northeastern states, especially during cool moist weather. The causal fungus, *Cladosporium cucumerinum*, produces small, circular, halo-bordered, water-soaked lesions on leaves and stems. The greatest damage is to the fruit and appears as sunken, dark-brown spots on immature fruit and as rough corky lesions on mature fruit. The fungus survives in plant refuse and on the seed. Two-year crop rotation, seed treatment, and well-drained soil are important in control of scab. Control is achieved by planting resistant varieties, such as the slicing variety Highmoor or the pickling varieties Wisconsin SR 6 and SMR 12.

Viral diseases. Cucumber mosaic is the most destructive disease caused by a virus. The pathogen causes mottling of terminal leaves and dwarfing of vines, as well as mottling and varying malformations of the fruits. Plants can be infected at any growth stage. The virus is harbored in a number of perennial hosts and is spread by aphids. The best means of control is provided by planting resistant varieties, such as the slicing varieties Burpee Hybrid, Niagara, Ohio MR 200, Sensation Hybrid, Shamrock, and Surecrop Hybrid, and pickling varieties Ohio MR 17, Ohio MR 25, Wisconsin SMR 12, and Yorkstate.

See HEMIPTERA; PLANT PATHOLOGY; PLANT VIRUSES AND VIROIDS.

Frank L. Caruso

Culture

The cultivation of cells in the laboratory. Bacteria and yeasts may be grown suspended in a liquid medium or as colonies on a solid medium (usually solidified by 1.5–2.5% agar); molds grow on moist surfaces; and animal and plant cells (tissue cultures) usually adhere to the glass or plastic beneath a liquid medium. Cultures must provide sources of energy and raw material for biosynthesis, as well as a suitable physical environment.

The materials supplied determine which organisms can grow out from a mixed inoculum. Some bacteria (prototrophic) can produce all their constituents from a single organic carbon source; hence they can grow on a simple medium (including K^+, NH_3, Mg^{2+}, Fe^{3+}, PO_4^{3-}, and SO_4^{2-}). Other cells (auxotrophic) lack various biosynthetic pathways and hence require various amino acids, nuclei acid bases, and vitamins. Obligatory or facultative anaerobes grow in the absence of O_2; many cells require elevated CO_2. Cultures isolated from nature are usually mixed; pure cultures are best obtained by subculturing single colonies. Viruses are often grown in cultures of a host cell, and may be isolated as plaques in a continuous lawn of those cells.

In diagnostic bacteriology, species are ordinarily identified by their ability to grow on various selective media and by the characteristic appearance of their colonies on test media. These tests include hemolytic reactions with sheep or other red cells embedded in the solid medium. Indicator dyes in the medium are used to detect colonies that ferment various sugars.

In ordinary cultures the cells are at all possible stages in their division cycle, and the composition of the medium changes continually as a result of their metabolism (until growth ceases, in the stationary phase of the culture). On transfer of an inoculum to fresh medium, there may be a lag phase, without multiplication, followed by a phase of exponential growth. Synchronous cultures are achieved by blocking growth or harvesting cells at a specific stage; the cells then divide in synchrony for several generations. In continuous cultures, fresh medium flows into the vessel and full-grown culture flows at the same rate (such as in a chemostat); the cells are therefore harvested from a medium of constant composition. *See* BACTERIAL GROWTH; CHEMOSTAT.

Laboratory cultures are often made in small flasks, test tubes, or covered flat dishes (petri dishes). Industrial cultures for antibiotics or other microbial products are usually in fermentors of 10,000 gallons (37,850 liters) or more. The cells may be separated from the culture fluid by centrifugation or filtration.

Bernard D. Davis

Technique. Specific procedures are employed for isolation, cultivation, and manipulation of microorganisms, including viruses and rickettsia, and for

propagation of plant and animal cells and tissues. A relatively minute number of cells, the inoculum, is introduced into a sterilized nutrient environment, the medium. The culture medium in a suitable vessel is protected by cotton plugs or loose-fitting covers with overlapping edges so as to allow diffusion of air, yet prevent access of contaminating organisms from the air or from unsterilized surfaces. The transfer, or inoculation, usually is done with the end of a flamed, then cooled, platinum wire. Sterile swabs may also be used and, in the case of liquid inoculum, sterile pipets.

The aqueous solution of nutrients may be left as a liquid medium or may be solidified by incorporation of a nutritionally inert substance, most commonly agar or silica gel. Special gas requirements may be provided in culture vessels closed to the atmosphere, as for anaerobic organisms. Inoculated vessels are held at a desired constant temperature in an incubator or water bath. Liquid culture media may be mechanically agitated during incubation. Maximal growth, which is visible as a turbidity or as masses of cells, is usually attained within a few days, although some organisms may require weeks to reach this stage. *See* CHEMOSTAT; EMBRYONATED EGG CULTURE; TISSUE CULTURE.

Media. Combinations of many nutrients are used for the growth of microorganisms as well as plant and animal tissues. Water is the chief constituent of most media. A medium may consist of natural materials, such as enzymatic digests or extracts of yeast or beef. Milk, potato slices, and chick embryos are other common examples. Artificial media are prepared by mixing various ingredients according to particular formulas. A complex medium contains at least one crude ingredient derived from a natural material, hence of unknown chemical composition. A chemically defined, or synthetic, medium is one in which the chemical structure and amount of each component are known.

All media furnish minerals, sources of nitrogen, carbon, and energy, and sometimes vitamins or other growth factors. An ingredient furnishing carbon—usually the chief source—is called the substrate for the utilizing organism.

Selective medium. Usually this medium contains an individual organic compound as the sole source of carbon, nitrogen, or sulfur for growth of an organism. The organism oxidizes the carbon compound, thereby obtaining energy for vital processes. Strictly inorganic media are utilized by some bacteria, the chemoautotrophs. They obtain carbon exclusively from carbon dioxide and energy from the oxidation of ammonia, nitrate, hydrogen gas, or reduced, inorganic, sulfur compounds. Algae and a few bacterial types are photoautotrophs, securing their energy from light.

An agar medium of a given composition permits selection, through development of colonies, or organisms capable of utilizing the specific sources of carbon, energy, nitrogen, or sulfur available. A colony on such a medium is prima-facie evidence that such organisms have been selected, but this must be confirmed by restreaking (subculturing on agar medium) on a medium of the same composition.

Enrichment media. These are liquid media of a given composition which permit preferential emergence of certain organisms that initially may have made up a relatively minute proportion of a mixed inoculum. This enrichment of the population in the desired type expedites isolation by direct plating on the corresponding selective solid medium. Because of the numerous subtle, competitive forces, enrichment cultures yield only a relatively small proportion, that is, the fastest growers, of all the organisms potentially capable of developing in a given enrichment medium.

Indicator media. These media, usually solid, contain substances capable of changing color in the vicinity of the colony which has effected a particular chemical change, such as fermenting a certain sugar. Thus, organisms with distinctive types of metabolism may be readily detected by the color change in the medium, even when the surface of the medium contains many other colonies. Indicators may be used in liquid media to characterize a pure culture of organisms.

Elective culture. No single culture medium or environment will support growth of more than a small proportion of the known types of microorganisms. Each organism has characteristic nutritional and environmental requirements. Thus, in a culture composed of more than one type of organism, the medium and conditions are selective for only one type of organism present and the other types are suppressed. These selective, or elective, culture principles and techniques are invaluable for the detection and isolation of specified organisms from natural sources. Elective culture techniques are more effective for obtaining bacteria from natural sources than for obtaining fungi, actinomycetes, and other groups; the faster growth rates, larger numbers, and greater variety of bacteria usually confer a competitive advantage upon them. In general, elective culture procedures are applicable to organisms capable of utilizing particular sources of energy, carbon, nitrogen, and other nutrients in a given physicochemical environment. There is no principle whereby elective cultures may be employed to obtain selectively organisms capable of producing particular compounds. Basically, two procedures are available, direct isolation and liquid enrichment.

Direct isolation. This requires a solid medium, usually agar, in petri plates. The surface is lightly sprinkled with finely divided particles of soil or other inoculum material. Only organisms favored by that medium and those conditions will produce colonies of appreciable size; these may then be purified. Literally thousands of other kinds of bacteria fail to develop, although they might develop in other selective media. Thus, the direct isolation technique is a means of selecting certain minority members from natural, mixed microbial populations. Since most of the colonies are on a solid medium, develop independently, and are physically separate, competition rarely is a factor determining the emergence of

the various colonies. Although certain fast-growing and characteristically spreading colonies may tend to overgrow neighboring colonies, the chief advantage of the procedure is that slow-growing bacteria have a chance to reveal themselves. This technique is, therefore, the best for obtaining the largest variety of bacteria capable of developing on a given medium. Colonies from a inoculum appearing under any one environmental condition are a small representation of a wider spectrum of bacteria capable of developing. For example, two sets of plates incubated, respectively, in the presence and absence of oxygen will yield different organisms, as will plates adjusted to different hydrogen ion concentrations (pH) and plates incubated at various temperatures.

Elective media containing common organic compounds, such as carbohydrates or amino acids, will yield large numbers of different organisms. Similarly, most such organisms are capable of utilizing a number of different but related compounds. However, certain substrates and conditions are highly specific for a few organisms or even for a single kind of organism. Occasionally, growth of a colony occurs at the expense of degradation of biosynthetic products or accessory growth factors which are liberated from one or more adjacent colonies and diffuse into the medium.

Although the direct-isolation method does not permit preferential development of colonies capable of forming specific products, such colonies as may be desired may be detected among those on an agar plate. Thus, acid producers may be distinguished by a zone of color change of an indicator in the medium or by dissolution of powdered calcium carbonate suspended in the medium. Clear zones result from the destruction of other insoluble substrates, such as starch, proteins, and lipids, by colonies producing the corresponding enzymes. Other metabolic products may be detected by flooding or spraying colony-containing plates with chemical or biological reagents, giving a distinctive reaction with the desired product.

The specificity can be enhanced by incorporation of substances that selectively inhibit growth of certain kinds of organisms, especially of those in the majority. For example, in a medium adjusted to pH 4.0–4.5 most bacteria and actinomycetes are inhibited, and thus filamentous fungi and yeasts may be selectively obtained. Other chemicals are used to inhibit undesirable organisms in specimens of body fluids, thereby facilitating isolation of the pathogenic bacteria in the minority.

Liquid enrichment. A small amount of soil or other natural material is added to a prescribed fluid medium which then is incubated. After appreciable growth of bacteria occurs, a drop of the culture is used to inoculate a second flask of the same medium. Organisms originally representing small minorities are thus preferentially selected by the medium and the conditions, and by the time turbidity develops in the secondary culture, they represent practically the majority populations. The chief disadvantage of this procedure is that only a small proportion of all the bacterial types potentially capable of developing in a given medium will be found. Potentially competent organisms are in direct competition, and the specific conditions usually select relatively few of the fastest growers. Liquid-enrichment cultures also contain organisms unable to attack the original substrate; these scavenge products of degradation and biosynthesis, as well as living and dead cells, of the primary organisms.

Pure culture. A pure culture contains cells of one kind, all progeny of a single cell. The concept of a pure culture is an important one in microbiology because most considerations of microorganisms require dealing with only one kind of a cell at a time. Many techniques have been evolved for the isolation and maintenance of pure cultures.

Isolation of aerobic bacteria. Isolation of pure cultures may be achieved by a variety of methods, the main differences being in isolation of aerobic versus anaerobic bacteria.

1. *Pour-plate procedures.* These involve a thorough distribution of a microbial mixture in cooled, melted agar medium. After solidification of the medium, individual cells are embedded and proliferate to form a colony. Purity of the colony cannot be taken for granted since it may have arisen from two adjacent cells (mixed colony); so the colony is dispersed in sterile water and the pour-plate procedure repeated. Well-isolated colonies in the second plating ordinarily can be shown to be pure. This method is based on the use of a homogeneous suspension of individual cells. The great majority of the colonies on secondary plates should be similar and, seen microscopically, the cells of such colonies should be one kind. However, certain organisms mutate characteristically, producing different colony types (dissociation).

2. *Streak plates.* The surface of the solid medium in a petri dish is streaked with an inoculating wire or glass rod in a continuous movement so that most of the surface is covered. The later parts of the streak should not intersect with the earlier parts. In later portions of the streak, the separate colonies are the outgrowth of individual cells. The chief advantage of this method is isolation of majority members of a mixed population.

3. *Roll-tube technique.* This technique, employed chiefly in tissue culture, achieves a homogeneous gaseous and nutrient environment by the elimination of diffusion gradients. During incubation the tubes are held in a wheellike instrument at an angle of about 15° from the horizontal. The wheel rotates vertically about once every 2 min. Constant liquid film formation over the inner surface of the tube facilitates oxygen transfer to the medium.

4. *Submerged culture.* In this method the microorganisms are grown in a liquid medium subjected to continuous, vigorous agitation. A homogeneous cellular suspension results and all cells in the medium are simultaneously provided with nearly optimal rates of diffusion of oxygen and nutrients. The mechanical agitation may be achieved by continuous bubbling of sterile air through the medium, by a

propeller system, by combination of both, or by continuous rotary or reciprocating shaking machines.

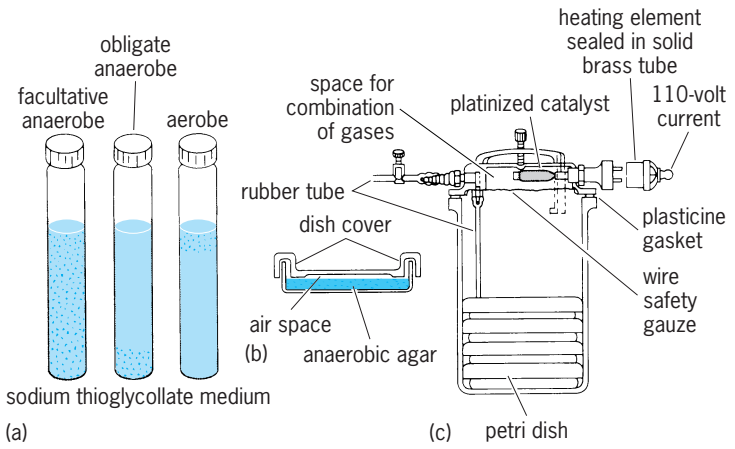
Isolation of anaerobic bacteria. Special cultural conditions may be provided for anaerobes, organisms that do not require oxygen or that grow only in the complete absence of oxygen. Fluid thioglycollate medium, the Brewer anaerobic petri dish, and the Brewer anaerobic jar are some of the methods used (see **illus.**). Fluid thioglycollate medium contains sodium thioglycollate, a reducing compound which lowers the oxygen content of the medium. In the anaerobic petri dish, the thioglycollate agar medium and the petri dish cover, with its restricted air space and close fit to the medium surface, give anaerobic conditions. The anaerobic jar is operated by replacing the air in the jar with hydrogen or illuminating gas. A platinized catalyst heated by electricity causes the oxygen remaining in the jar to combine with hydrogen to form water or, in the case of illuminating gas, water and carbon dioxide. In this way free oxygen is removed from the jar.

1. *Shake culture.* This is a method for isolating anaerobic bacteria. Various dilutions of an inoculum are throughly mixed with cooled, melted agar medium occupying about two-thirds the volume of test tubes. After solidification a 1-in. (2.5-cm) layer of melted vaspar, a 50:50 mixture of paraffin wax and petrolatum, is added. When the mixture has hardened, the culture is sealed against diffusion of oxygen during incubation. In tubes containing higher dilutions of inoculum, separated discrete colonies develop in different positions in the agar. With soft glass tubes, the glass is cut and the agar column broken in the region of a desired colony, which is then isolated. Sometimes the seal is first removed by warming that area of the tube; the column of agar is then slowly ejected into a sterile container by compressed air carried to the bottom of the tube through a fine capillary tube introduced between the agar and the glass. Colonies are reached by slicing the agar with a sterile scalpel.

2. *Anaerobic plates.* Pour or streak plates are incubated in a closed vessel in which the air is replaced by an inert oxygen-free gas such as helium or nitrogen. Alternatively, the oxygen in a closed vessel is absorbed chemically by mixing pyrogallol and alkali. Specially designed petri dishes permit oxygen absorption in individual dishes.

Single-cell isolation. Single-cell isolation employs microscopic techniques to verify the presence of a single cell, which is then used to initiate a culture.

1. *Little-drop technique.* Minute drops of dilutions of a cellular suspension are examined microscopically on a glass slide. A drop containing a single cell is transferred with a capillary pipet to an appropriate culture medium. This procedure is best suited for isolation of large cells, such as yeasts and mold spores, which can be located quickly with the high-dry lens of the compound microscope. Evaporation problems and the difficulty of obtaining drops small enough to examine with the oil immersion lens limit this procedure for bacteria. A variation of this procedure was originated by J. Lederberg. Little drops of a diluted cellular suspension are deposited on a slide covered with a thin film of oil. The Lederberg technique, by permitting rapid isolation of single bacterial cells, obviates to some extent the problems enumerated before.



Some techniques for the cultivation of anaerobic bacteria. (*a*) Fluid thioglycollate medium showing aerobic, anaerobic, and facultative anaerobic growth. (*b*) Brewer anaerobic petri dish. (c) Brewer anaerobic jar. (*Baltimore Biological Laboratory, Inc.*)

2. *Micromanipulation.* The disadvantages of the little-drop technique for the isolation of single bacterial cells are eliminated with this technique. A micromanipulator permits study of drops in a moist chamber on the microscope stage, reducing evaporation difficulties and enabling the use of the oil immersion lens. The manipulations are performed by remotely controlled capillary micropipets activated by air pressure. *See* MICROMANIPULATION.

Maintenance. This is concerned with the preservation of culture viability and is accomplished by practices that reduce the basal metabolism of the cells, thereby enhancing longevity and eliminating the need for frequent transfers.

1. *Lyophilization.* This procedure, also called freeze-drying, is the best method for preserving culture viability. Freshly grown cells are suspended, in a glass tube, in a small volume of a sterile fluid, such as blood serum or skim milk. A small amount of a reducing agent such as ascorbic acid may help prolong the viability of the culture. The suspension is quickly frozen at approximately −94°F (−70°C) by immersion in a dry ice–acetone mixture. The frozen material is then dehydrated, without melting, under high vacuum, or the gas space may be filled with nitrogen or helium prior to sealing. Storage in the absence of oxygen may preserve the viability of a culture for years, even without refrigeration. The mortality, during lyophilization and storage, varies with the culture and the suspending fluid. Lyophilized cells are rejuvenated by inoculation into fresh nutrient medium.

2. *Slants.* Slants are the most convenient method of maintenance for periods of 1–2 months. Cotton-plugged tubes containing an agar medium are slanted after sterilization until the agar solidifies. The inoculum is streaked on the agar surface. After maximal growth at the appropriate temperature, the slant culture is stored in a refrigerator and used repeatedly as

inoculum. With long storage periods, the tube should be sealed to retard desiccation. Viability of slant cultures not used frequently may be prolonged by filling the tube with sterile mineral oil to about 1 in. (2.5 cm) above the agar. This retards water loss from and oxygen access to the culture.

3. *Stabs.* Stabs are used to maintain anaerobic bacteria. A tube about two-thirds filled with agar medium is pierced vertically for about 2 in. (5 cm) with an inoculating wire. Growth develops in the anaerobic depths of the agar along the stab.

4. *Strict anaerobe maintenance.* Strict anaerobes, in liquid or solid media, may be maintained in an atmosphere from which the oxygen has been absorbed by alkaline pyrogallol. Also, the air may be replaced by an inert gas, such as helium or nitrogen. The most convenient procedure employs liquid media containing approximately 0.1% of a reducing agent, such as sodium thioglycollate, to keep the oxidation-reduction potential low enough for growth of anaerobes and to protect it from oxygen diffusing from the air. Special vessels or atmospheres are not required, but fully developed cultures are sealed against oxygen and refrigerated.

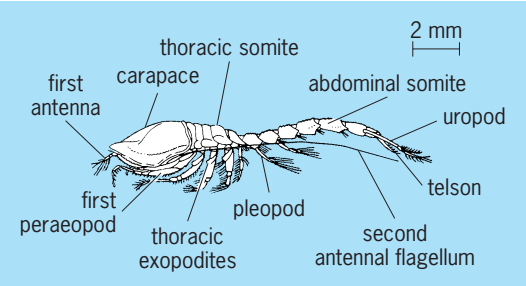
5. *Use of sterile soil.* Sterile soil may be used to carry stock cultures of sporeforming organisms. Autoclaved soil supports limited growth of the organism, which then sporulates. The spores remain viable indefinitely when kept refrigerated.

Jackson W. Foster; R. E. Kallio

Cumacea

A crustacean within the Malacostraca: Peracarida. All have a well-developed carapace which is fused dorsally with at least the first three thoracic somites and which overhangs the sides to enclose a branchial cavity. The telson may be free or may appear lacking because it is fused with the last abdominal somite. Eyes are sessile and, except for a few genera, are coalesced in a single median organ or wholly wanting. There are eight pairs of thoracic appendages, the first three modified as maxillipeds and the last five as peraeopods. The first maxillipeds bear an epipodite forming a respiratory organ with a forward-directed siphon and a posterior part carrying branchial lobes. Natatory exopodites may be present on the third maxillipeds and the first four pairs of peraeopods or some of them. The uropods are styliform. There is no free larval stage. In the few species whose life history has been investigated there are five to seven molts after release from the brood pouch.

There is always some sexual dimorphism in adults, but immature males resemble females until the later stages. The second antennae are always vestigial in females and nearly always well developed with a long flagellum in males (see **illus.**). Females often have a reduced number of thoracic exopodites. Pleopods are never present in the female but there may be one to five pairs in the male. The form and sculpturing of the body and the number of spines and setae on the



Typical adult male cumacean.

appendages often differ between the sexes. Oostegites are present on the third to sixth thoracic limbs of the female.

All known species have a characteristic shape owing to the inflated carapace and slender abdomen. Most are small, 0.04–0.48 in. (1–12 mm) in length; the largest reach 1.4 in. (35 mm). The order is rather uniform and classification is based on trivial characters. Seven families are recognized. Affinities are probably closer to the Mysidacea than to the Tanaidacea. A few fossils are now known.

Cumacea are found in all seas and in some estuaries and areas of brackish water, including the Caspian Sea. Most inhabit water shallower than 650 ft (200 m), but many occur in great depths. They normally burrow in sand or mud with the front protruding. Coastal species, and especially the more active adult males, may often be captured above the bottom. They feed on organic detritus or very small organisms and may themselves form part of the diet of various fishes. About 740 species have been described and placed in 82 genera. Many more remain to be found, as they have been poorly investigated in most areas. *See* MALACOSTRACA; SEXUAL DIMORPHISM.

Norman S. Jones

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Cumin

Cuminum cyminum, a plant for which the whole or ground dried ripe fruit, commonly called seed, is a popular spice. It is a major ingredient in both chili

powder and curry, and is added to meat sauces, rice, bread, pickles, soups, and other foods.

Roman caraway is another common name for this member of the parsley family (Apiaceae). The only species in its genus, cumin exhibits a variety of plant types depending on the seed source. A small annual herb about 1–2 ft (0.3–0.6 m) tall, cumin grows upright as a single slender stem with many branches. The deep green leaves are 0.5–2 in. (1.3–5 cm) long, thin, and finely divided. *See* APIALES.

The strong, pungent green-spicy odor and flavor of cumin is attributable largely to cuminaldehyde, the main constituent of the essential oil, and other aldehydes.

This herb is native to the Mediterranean region. Presently cumin is commercially grown in Iran, southern Russia, China, India, Morocco, and Turkey. The three major types of cumin seed, Iranian, Indian, and Middle Eastern, vary in seed color, essential oil quantity, and flavor. Cumin does best in a mild growing climate, free from hot dry periods, on well-drained sandy loam soil. In 90–120 days the plant matures; the small fragile plant is generally hand-harvested. Commonly cumin is planted at 30–35 lb/acre (33.7–39.3 kg/ha), and yields 400–500 lb seed/acre (450–560 kg/ha). Upon distillation cumin seed yields 2.5–5% essential oil, used in both perfumery and flavoring liqueurs. Cumin is also used medicinally. *See* SPICE AND FLAVORING.

Seth Kirby

Cummingtonite

An amphibole (a double-chain silicate mineral) with the idealized chemical formula $Mg_7Si_8O_{22}(OH)_2$ that crystallizes with monoclinic symmetry. Naturally occurring samples generally are solid solutions between $Mg_7Si_8O_{22}(OH)_2$ and the corresponding iron (Fe)-end member with the general formula $(Mg,Fe)_7Si_8O_{22}(OH)_2$. The name cummingtonite (derived from the location Cummington, Massachusetts) is applied to all solid solutions with $Mg/(Mg + Fe^{2+}) \geq 0.5$, whereas those with $Mg/(Mg + Fe^{2+}) < 0.5$ are termed grunerite. Up to a total of 1.0 (Ca + Na) atom per formula unit may also be present in cummingtonite.

The amphibole anthophyllite also has the idealized formula $Mg_7Si_8O_{22}(OH)_2$, but crystallizes with orthorhombic symmetry. The conditions that favor formation of one amphibole over the other are not fully understood, but both are known to coexist in the same rock. It has been proposed that cummingtonite is stable at higher temperature, and inverts, on cooling, to anthophyllite by a process that involves unmixing of a more calcium-rich amphibole (tremolite) and the formation of an intermediate metastable monoclinic form of cummingtonite.

Cummingtonite commonly occurs as aggregates of fibrous crystals, often in radiating clusters. It is transparent to translucent, varies in color from white to green to brown, and may be pale to dark depending primarily on the iron content. Hardness is 5–6 on Mohs scale; density is 3.1–3.6 g/cm^3 (1.8–2.1 oz/in.^{3}), increasing with increasing iron content. Cummingtonite has distinct, prismatic (110) cleavage that produces cross sections with approximately 124° and 56° angles, which are definitive for amphibole group minerals. Chemical, optical, or x-ray techniques usually are necessary to unambiguously distinguish cummingtonite from other amphibole minerals.

Cummingtonite is generally considered to be a metamorphic mineral, but it has been found in silicic volcanic rocks and, rarely, plutonic igneous rocks. It occurs in a variety of metamorphic rock types (amphibolite, schist, gneiss, granulite) that have undergone medium- to high-grade metamorphism. It commonly occurs in the calcium- and aluminum-poor environment of metamorphosed iron formation. It can also be a constituent of metamorphosed mafic and ultramafic igneous rocks, where it may coexist with other amphibole minerals such as hornblende, tremolite, and anthophyllite. With increasing intensity of metamorphism, cummingtonite is commonly replaced by pyroxene-bearing mineral assemblages. *See* AMPHIBOLE; HORNBLENDE; IGNEOUS ROCKS; METAMORPHIC ROCKS; TREMOLITE.

Robert K. Popp

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Curie temperature

The critical or ordering temperature for a ferromagnetic or a ferrimagnetic material. The Curie temperature (T_C) is the temperature below which there is a spontaneous magnetization M in the absence of an externally applied magnetic field, and above which the material is paramagnetic. In the disordered state above the Curie temperature, thermal energy overrides any interactions between the local magnetic moments of ions. Below the Curie temperature, these interactions are predominant and cause the local moments to order or align so that there is a net spontaneous magnetization.

In the ferromagnetic case, as temperature T increase from absolute zero, the spontaneous magnetization decreases from M_0, its value at $T = 0$. At first this occurs gradually, then with increasing rapidity until the magnetization disappears at the Curie temperature (see **illus.**). Many physical properties show an anomaly or a change in behavior at the Curie temperature. There is a peak in specific heat and in magnetic permeability, and there are changes in the behavior of such properties as electrical

Ordering temperatures of magnetic materials*

Ferromagnets	Curie temperature (T_C), °F (K)	Ferrimagnets	Curie temperature (T_C), °F (K)	Antiferromagnets	Néel temperature (T_N), °F (K)
Co	2039 (1388)	Fe_3O_4	1085 (858)	NiO	620 (600)
Fe	1418 (1043)	$MnFe_2O_4$	572 (573)	Cr	100 (311)
$CrBr_3$	−393 (37)	$Y_3Fe_5O_{12}$	548 (560)	$FeCl_2$	−417.0 (23.7)
$GdCl_3$	−455.7 (2.2)				

*From F. Keffer, *Handbuch der Physik*, vol. 18, pt. 2, Springer, 1966.

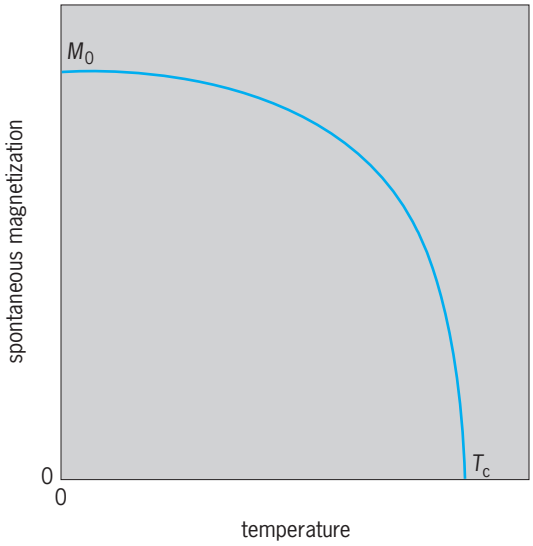
resistivity, elastic properties, and thermal expansivity. In ferrimagnetic materials the course of magnetization with temperature may be more complicated, but the spontaneous magnetization disappears at the Curie temperature. The Curie temperature can be determined from magnetization versus temperature measurements or from the related anomalies in other physical properties. The Curie temperature can be altered by changes in composition, pressure, and other thermodynamic parameters.

Well above the Curie temperature the magnetic susceptibility follows the Curie-Weiss law, and in fact the characteristic temperature θ in that the law is called the paramagnetic Curie temperature. Generally θ is slightly above T_C. The susceptibility deviates from Curie-Weiss behavior as the temperature approaches T_C and θ. *See* CURIE-WEISS LAW; FERRIMAGNETISM; FERROMAGNETISM; MAGNETIC SUSCEPTIBILITY; MAGNETIZATION; PARAMAGNETISM.

In antiferromagnetic materials the corresponding ordering temperature is termed the Néel temperature (T_N). Below the Néel temperature the magnetic sublattices have a spontaneous magnetization, though the net magnetization of the material is zero. Above the Néel temperature the material is paramagnetic. *See* ANTIFERROMAGNETISM.

The ordering temperatures for magnetic materials vary widely (see **table**). The ordering temperature for ferroelectrics is also termed the Curie temperature, below which the material shows a spontaneous electric moment. *See* FERROELECTRICS; PYROELECTRICITY.

J. F. Dillon, Jr.



Variation of spontaneous magnetization of a ferromagnetic material with temperature.

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Curie-Weiss law

A relation between magnetic or electric susceptibilities and the absolute temperature which is followed by ferromagnets, antiferromagnets, nonpolar ferroelectrics and antiferroelectrics, and some paramagnets. The Curie-Weiss law is usually written as the equation below, where χ is the susceptibility, C is a

$$\chi = C/(T - \theta)$$

constant for each material, T is the absolute temperature, and θ is called the Curie temperature. Antiferromagnets and antiferroelectrics have a negative Curie temperature. The Curie-Weiss law refers to magnetic and electric behavior above the transition temperature of the material in question. It is not always precisely followed, and it breaks down in the region very close to the transition temperature. Often the susceptibility will behave according to a Curie-Weiss law in different temperature ranges with different values of C and θ. *See* CURIE TEMPERATURE; ELECTRIC SUSCEPTIBILITY; MAGNETIC SUSCEPTIBILITY.

The Curie-Weiss behavior is usually a result of an interaction between neighboring atoms which tends to make the permanent atomic magnetic (or induced electric) dipoles point in the same direction. The strength of this interaction determines the Curie temperature θ. In the magnetic case the interaction is caused by Heisenberg exchange coupling.

In many paramagnetic salts, especially those of the iron group, the Curie-Weiss behavior is due to crystalline distortions and their effect upon the atomic orbital angular momenta which contribute to the magnetization. For further discussion and derivation of the Curie-Weiss law. *See* ANTIFERROMAGNETISM; FERROELECTRICS; FERROMAGNETISM.

Elihu Abrahams; Frederic Keffer

Curium

A chemical element, Cm, in the actinide series, with an atomic number of 96. Curium does not exist in the terrestrial environment, but may be produced artificially. The chemical properties of curium are so similar to those of the typical rare earths that, if it were not for its radioactivity, it might easily be mistaken for one of these elements. The known isotopes of curium include mass numbers 238–250. The isotope ^{244}Cm is of particular interest because of its potential use as a compact, thermoelectric power source, through conversion to electrical power of the heat generated by nuclear decay. *See* PERIODIC TABLE.

1																	18
1 H	2											13	14	15	16	17	2 He
3 Li	4 Be											5 B	6 C	7 N	8 O	9 F	10 Ne
11 Na	12 Mg	3	4	5	6	7	8	9	10	11	12	13 Al	14 Si	15 P	16 S	17 Cl	18 Ar
19 K	20 Ca	21 Sc	22 Ti	23 V	24 Cr	25 Mn	26 Fe	27 Co	28 Ni	29 Cu	30 Zn	31 Ga	32 Ge	33 As	34 Se	35 Br	36 Kr
37 Rb	38 Sr	39 Y	40 Zr	41 Nb	42 Mo	43 Tc	44 Ru	45 Rh	46 Pd	47 Ag	48 Cd	49 In	50 Sn	51 Sb	52 Te	53 I	54 Xe
55 Cs	56 Ba	71 Lu	72 Hf	73 Ta	74 W	75 Re	76 Os	77 Ir	78 Pt	79 Au	80 Hg	81 Tl	82 Pb	83 Bi	84 Po	85 At	86 Rn
87 Fr	88 Ra	103 Lr	104 Rf	105 Db	106 Sg	107 Bh	108 Hs	109 Mt	110 Ds	111 Rg	112	113					

lanthanide series	57 La	58 Ce	59 Pr	60 Nd	61 Pm	62 Sm	63 Eu	64 Gd	65 Tb	66 Dy	67 Ho	68 Er	69 Tm	70 Yb
actinide series	89 Ac	90 Th	91 Pa	92 U	93 Np	94 Pu	95 Am	96 Cm	97 Bk	98 Cf	99 Es	100 Fm	101 Md	102 No

Metallic curium may be produced by the reduction of curium trifluoride with barium vapor. The metal has a silvery luster, tarnishes in air, and has a specific gravity of 13.5. The melting point has been determined as $2444 \pm 72^\circ$F ($1340 \pm 40^\circ$C). The metal dissolves readily in common mineral acids with the formation of the tripositive ion.

A number of solid compounds of curium have been prepared and their structures determined by x-ray diffraction. These include CmF_4, CmF_3, CmC_3, $CmBr_3$, CmI_3, Cm_2O_3, and CmO_2. Isostructural analogs of the compounds of curium are observed in the lanthanide series of elements. *See* ACTINIDE ELEMENTS; TRANSURANIUM ELEMENTS. Glen T. Seaborg

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Currant

A fruit (berry) in the genus *Ribes* in the family Saxifragaceae. Cultivated black and red currants and gooseberries all belong to this genus. *Ribes* species having prickles or spines are called gooseberries, and those that do not are called currants. The berries are produced in clusters on bushes, and cultivars ripen in midsummer (**Fig. 1**).

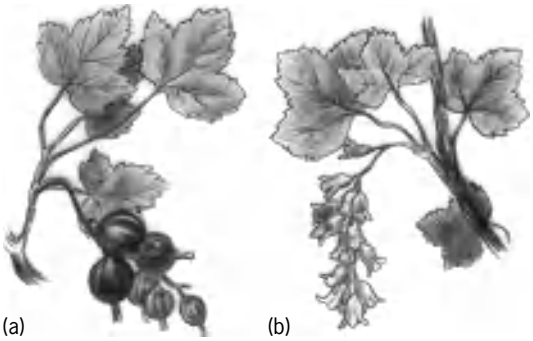

Fig. 1. Black-currant (*Ribes nigrum*). (*a*) Branch with cluster of fruits. (*b*) Branch with flowers.

Varieties. *Ribes* are not widely grown commercially for fruit in the United States, but red currants and gooseberries are popular in home gardens for use in jellies and pies. Several *Ribes* species, particularly golden currant (*R. aureum*) and fuchsia-flowered gooseberry (*R. speciosum*), are used as shrubs in landscaping. In the United States important red currant cultivars include Cherry, Minnesota 71, Perfection, Red Lake, White Imperial, and Wilder; popular gooseberry cultivars include Fredonia, Pixwell, Poorman, Red Jacket, and Welcome. In Europe cultivars of the black currant (*R. nigrum*) are extensively grown commercially for juice concentrate. Wild *Ribes* species occur widely in the United States, and desirable, edible types are sometimes gathered during the season when the fruit is locally abundant. *See* GOOSEBERRY.

Currant and gooseberry diseases. Many of the same diseases affect both currants and gooseberries. Several virus diseases occur in Europe in these crops, reducing yield and fruit quality, often without showing symptoms on the leaves. In Great Britain and several other European countries where *Ribes* species are important crops, local growers can purchase virus-tested, certified plants from nurseries, thus reducing the likelihood of planting virus-infected stock. Tomato ringspot virus, causing leaf-spotting patterns and reduced vigor in red currants in the United States, is spread in the soil by the American dagger nematode (*Xiphinema* sp.), and can be controlled by appropriate soil fumigation prior to planting.

Anthracnose is a leaf- and fruit-spotting fungus disease caused by *Pseudopeziza ribis* (**Fig. 2**). It may be serious in wet seasons, causing leaves to yellow and drop by midsummer. Control involves elimination of overwintered leaves that harbor the fungus; pruning to allow rapid drying of interior foliage; and use of approved fungicides, usually fixed coppers, at bloom and harvest.

Powdery mildew (caused by *Sphaerotheca morsuvae*) is a fungus disease that affects cultivated *Ribes*, particularly gooseberries, producing white, powdery coatings that later darken on young shoots, leaves, and fruits. Powdery mildew often stunts infected plants and renders the fruit visually unappealing. The disease is hard to control, but sanitation and spraying with approved fungicides just as the buds open in spring and during the growing season often help.

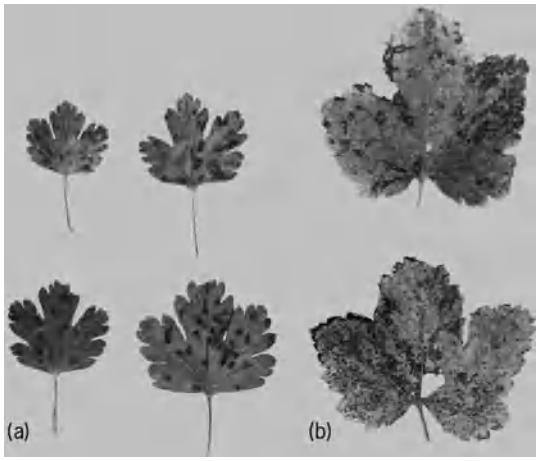

Fig. 2. Anthracnose disease on (*a*) gooseberry (small leaves) and (*b*) currant (large leaves).

Other important fungus diseases of *Ribes* include white pine blister rust (caused by *Cronartium ribicola*) on black currant, leaf spot (caused by *Mycosphaerella ribis*), and cane blight (caused by *Botryosphaeria ribis*). *See* PLANT PATHOLOGY; ROSALES.

Richard H. Converse

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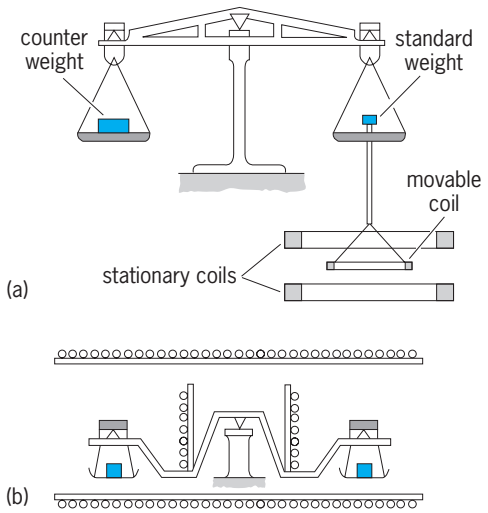
Current balance

An apparatus with which force is measured between current-carrying conductors. The purpose of the measurement was to establish the value of the ampere in terms of its International System (SI) definition; therefore the relative locations of the conductors had to be measured so that the force between them could be computed from their geometry. The difficulty of this measurement limited the accuracy of even the best instruments to a few parts in a million.



Current balances. (*a*) Rayleigh. (*b*) Pellat.

Therefore, current balances have now been superseded by other methods. *See* ELECTRICAL UNITS AND STANDARDS; PHYSICAL MEASUREMENT.

Operation. In a current balance, force is measured between fixed and movable coils (see **illus.**). If the coils carry the same current, Eq. (1) holds for a linear displacement, or Eq. (2) for an angular displacement, where F is the mechanical force or Γ is the mechanical torque, M is the coefficient of mutual inductance, and I is the current being measured. The coil construction must be such that $\partial M/\partial x$ or $\partial M/\partial\theta$ can be computed from dimensional measurements. The movable coil is attached to a balance, and the force between the coils is balanced by the gravitational force on a known mass.

$$F = I^2 \frac{\partial M}{\partial x} \tag{1}$$

$$\Gamma = I^2 \frac{\partial M}{\partial \theta} \tag{2}$$

Types. Three types of current balance have been used in ampere determinations. In two of these, the Rayleigh and the Ayrton-Jones balances, the fixed and movable coils are coaxial, and the motion is linear along the coil axis. In the third type, the Pellat balance, the axes of the fixed and movable coils are at right angles, and the movable coil turns about its center, the pivot point of the weighing balance.

In the Rayleigh balance the coils are circular and have a small square winding section. The fixed coils are located above and below, but coaxial with, the smaller, movable coil. The coil spacing is such that the force on the movable coil is maximum. Because of the small winding section, the ratio of radii of fixed and movable coils ratio can be calculated from the ratio of ampere-turns which produce cancellation of the magnetic flux densities at the center of the coil system. The radius of the movable coil is about half that of the fixed coils because for this ratio the exact dimensions of the winding cross section are of reduced importance.

In the Ayrton-Jones and Pellat balances, the fixed and movable coils are single-layer solenoids and the coil construction is such that the diameter and winding pitch can be measured accurately.

The current being measured was passed through a standard resistor, and the resulting voltage drop was compared to the emf of a standard cell. The resistor and cell were used to preserve the values of the electrical units in the intervals between SI ampere and ohm determinations. *See* CURRENT MEASUREMENT.

Bryan P. Kibble

Current comparator

A device that measures very accurately an unknown current in terms of a known one. Current comparators are based on Ampère's law, Eq. (1), where the

$$\sum I_p \cdot n_p = \oint \vec{H} \cdot d\vec{l} \tag{1}$$

left side of the equation is the sum of currents, each calculated as the current I_p in a conductor multiplied by the number of times n_p it threads a closed path in space (this product is the so-called ampere-turns or current linkage), and the right side is the integral of the magnetic field strength $\vec{H}$ that the currents produce around the entire path. If an instrument could detect when the result of the path integral was exactly zero, then (supposing that only two currents I_1 and I_2 thread the path n_1 and n_2 times, respectively) we would know that there was accurate ampere-turn balance, that is, Eq. (2) is valid. The shape and position of the current circuits is in principle immaterial. *See* AMPÈRE'S LAW; MAGNETISM.

$$I_1 \cdot n_1 = I_2 \cdot n_2 \tag{2}$$

Room-temperature current comparators. Ideal flux line–integral detectors do not exist. If alternating currents are being compared, detecting the voltage induced at the contiguous ends of a solenoid uniformly close-wound on the surface of a tube that follows the path, which is usually circular (**Fig. 1**), is an approximation which will be limited in accuracy by imperfections of the winding to a fraction of a percent. Sensitivity can be greatly enhanced if the tube contains a highly magnetically permeable material. If the currents are steady direct currents, the magnetic material and the coil will be operated as a fluxgate magnetometer. To attain a million times better accuracy than a percent requires in practice that the current circuits are so positioned and shaped that the resulting magnetic field follows the circular path as closely as possible. In room-temperature current comparators, this is accomplished by winding the current-carrying conductors on nested toroids that enclose the detector's path. Additionally, this path is contained within a toroidal shield of magnetically highly permeable material (Fig. 1). This suppresses all field components except those following the detector's toroidal path. *See* MAGNETOMETER.

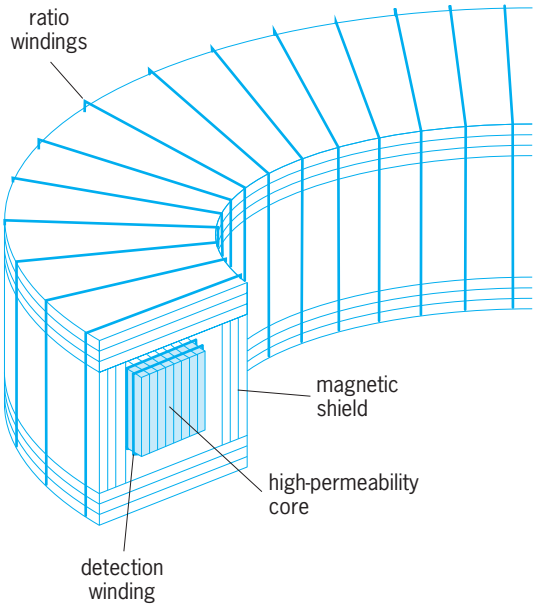


Fig. 1. Room-temperature alternating-current (AC) comparator.

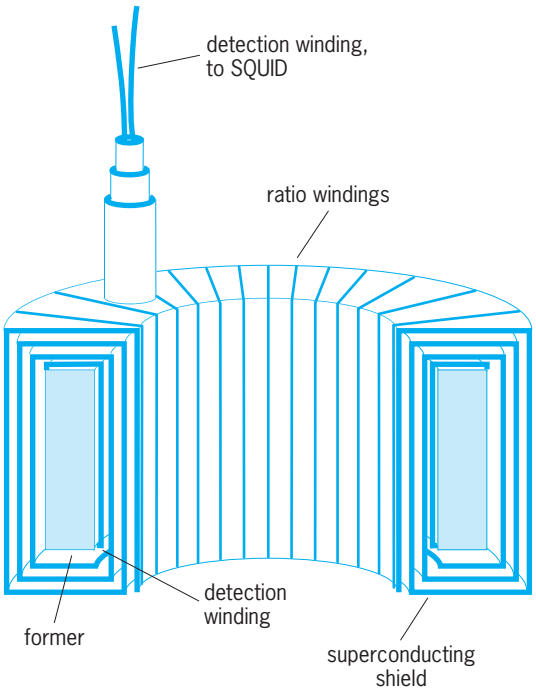


Fig. 2. Cryogenic current comparator.

Cryogenic current comparators. In superconducting cryogenic comparators (**Fig. 2**), which are principally used in national measurement laboratories in order to attain the highest possible level of demonstrable accuracy, flux shaping is achieved by making use of the Meissner effect. Several types of cryogenic comparators exist, but in all the current-carrying windings thread a toroidal superconducting shield, consisting of a tube folded back on itself, creating an overlap of several turns (Fig. 2). The Meissner effect ensures that the magnetic field inside the superconductor decreases exponentially with distance from the surface, rapidly becoming negligibly small, and consequently the flux line integral along a path well within the superconductor is also negligibly small. It follows from Ampère's law that the total current linkage enclosed by the path must also be negligible. As a consequence, screening currents flow on the surface of the superconductor, which oppose the current linkage imposed by any imbalance in the currents in the windings. On the surface that faces the windings, the distribution of the surface currents depends on the position of the current-carrying windings; but on the other side, well away from the ends, the current (and the field) distribution becomes very nearly independent of the position of windings. A detection winding in this region is therefore sensitive only to the total current linkage of the ratio windings. The perfection of the flux shaping is such that the pick-up coil need not be wound with as much precision as is needed for the room-temperature instrument. *See* MEISSNER EFFECT.

Flux detection in cryogenic current comparators is usually achieved by using a superconducting detection winding connected directly to the input of a SQUID. Compared to a room-temperature direct-current (DC) comparator, this has the advantage that no excitation winding is required, as a direct-current in the windings results in a DC signal at the output of the SQUID system. On the other hand, operation of a current comparator with a SQUID, which is very sensitive to electrical interference, requires the use of purpose-built low-noise electronics. A cryogenic comparator can achieve an accuracy exceeding 1 part in 10^9. *See* SQUID; SUPERCONDUCTING DEVICES.

In addition to imperfect flux shaping, another kind of error results from the electric fields associated with the potential drops down the resistive conductors carrying the currents. If the currents change in time, for example, if alternating currents are being compared, the time dependence of the electric fields results in displacement currents whose magnetic effects are indistinguishable from currents flowing in conductors. Fortunately, displacement currents are zero if steady direct currents are being compared, and their effect is very small if the frequency of the alternating current is only a few hundred hertz and the resistance of the windings is not unduly great. *See* DISPLACEMENT CURRENT.

Applications. Applications of current comparators range from the measurement of very large currents (of the order 1000 A for room-temperature comparators), to medium currents, such as used in the most accurate resistance measurements, and small currents (of the order 1 nA), such as generated by single-electron transport. *See* CURRENT MEASUREMENT; ELECTRICAL MEASUREMENTS; RESISTANCE MEASUREMENT.

Peter Kleinschmidt; Bryan P. Kibble

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Current density

A vector quantity equal in magnitude to the instantaneous rate of flow of electric charge, perpendicular to a chosen surface, per unit area per unit time. If a uniform wire of cross-sectional area A carries a current I, the current density J is I/A. The units of J in the International System (SI) are amperes per square meter.

The concept of current density is useful in treating the flow of electricity through a conductor of nonuniform cross section, and in electromagnetic theory. In terms of current density Ohm's law can be written $J = E/\rho$, where E is electric field strength and ρ is the local resistivity. Current density is related to the number per unit volume n, the charge e, and the average velocity υ of the effective charge carriers in a conductor by the formula $J = ne\upsilon$.

John W. Stewart

Current measurement

The measurement of the rate of passage of electric charges in a circuit. The unit of measurement, the ampere (A), is one of the base units of the International System of Units (SI). It is defined as that constant current which, if maintained in two straight parallel conductors of infinite length, of negligible circular cross section, and placed 1 m apart in vacuum, would produce between these conductors a force equal to 2×10^{-7} newton per meter of length.

Determination of the ampere. In order to establish an electrical unit in accordance with the SI definition, it is necessary to carry out an experimental determination. Obviously the ampere cannot be realized exactly as defined. Electromagnetic theory has to be used to relate a practical experiment to the definition. For many years the current balance was used, but its accuracy is limited to about 10 parts per million by the difficulties of the mechanical measurements to define coils and convection currents. *See* CURRENT BALANCE.

An alternative determination of the SI unit of current, devised by B. P. Kibble, depends on the availability of a satisfactory SI standard of resistance. The unit of resistance can be set up through the calculable capacitor, which is employed in several national standards laboratories. Through the concept of frequency (the reciprocal of time, another SI base quantity), the unit of impedance obtained from a calculable capacitor can be used to calibrate a resistor in SI units. In Kibble's virtual power balance, a coil is suspended partly in a strong permanent magnetic field. The experiment has two parts. First, a current I is passed through the coil and the resulting force F is measured by weighing. The current and force are proportional and related by a factor depending on the geometry of the apparatus. Second, the coil is moved at a controlled speed and the voltage induced in the coil is measured. The voltage V and the rate of displacement (dx/dt) are also proportional and depend on the same geometrical factor. The results of the two operations can be combined to eliminate the geometrical factor, giving the result shown in the equation below, where electric power = mechanical power. The result of the experiment, when combined with the SI unit of resistance (V/I), allows the ampere to be determined and all the electrical units to be set up in accordance with the SI definitions. *See* RESISTANCE MEASUREMENT.

$$V \times I = F \times \frac{dx}{dt}$$

Current or virtual power balances are not convenient for use as working standards. Since

January 1, 1990, working standards of voltage and resistance have provided the foundations of practical electrical measurements. The standard of voltage is based on the alternating-current (ac) Josephson effect, in which voltage is related to frequency. By international agreement the value 483 597.9 GHz/V for the Josephson constant is now used throughout the world. The working unit of resistance is maintained through the quantum Hall effect, with an agreed value of 25 812.807 ohms for the voltage-to-current ratio obtained under certain defined experimental conditions. These values have been chosen to provide the best known approximations to the SI units and have the advantage of reproducibility at the level of 1 part in 10^8. The working standard of current is derived from measurements of voltage across a known resistor. *See* ELECTRICAL UNITS AND STANDARDS; HALL EFFECT; JOSEPHSON EFFECT.

Direct and low-frequency currents. The moving-coil (d'Arsonval) meter measures direct currents (dc) from 10 microamperes to several amperes. The accuracy is likely to be a few percent of the full-scale indication, although precision instruments can achieve 0.1% or even better. Above 1 milliampere a shunt usually carries the major part of the current; only a small fraction is used to deflect the meter. Since the direction of deflection depends on the direction of the current, the d'Arsonval movement is suitable for use only with unidirectional currents. Rectifiers are used to obtain dc and drive the meter from an ac signal. The resulting combination is sensitive to the rectified mean value of the ac waveform.

In the moving-iron meter, two pieces of soft magnetic material, one fixed and one movable, are situated inside a single coil. When current flows, both pieces become magnetized in the same direction and accordingly repel each other. The moving piece is deflected against a spring or gravity restoring force, the displacement being indicated by a pointer. As the repulsive force is independent of current direction, the instrument responds to low-frequency ac as well as dc. The natural response of such a movement is to the root-mean-square (rms) value of the current. The accuracy of moving-iron meters is less than that of moving-coil types. *See* AMMETER.

Radio-frequency currents. For radio-frequency applications it is essential that the sensing element be small and simple to minimize inductive and capacitive effects. In a thermocouple meter the temperature rise of a short, straight heater wire is measured by a thermocouple and the corresponding current is indicated by a d'Arsonval movement. In a hot-wire ammeter the thermal expansion of a wire heated by the current is mechanically enhanced and used to deflect a pointer. Both instruments, based on heating effects, respond to the rms value of the current. Above 100 MHz, current measurements are not made directly, as the value of current is likely to change with position owing to reflections and standing waves. *See* MICROWAVE MEASUREMENTS; THERMOCOUPLE.

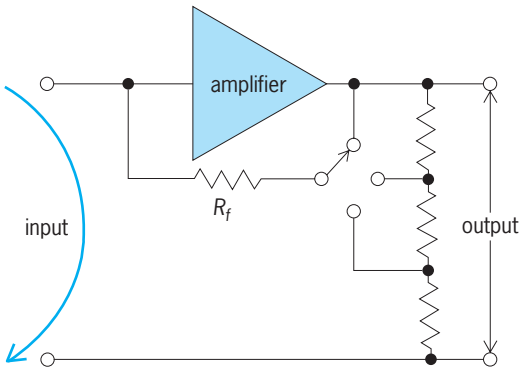
Large currents. Above 50 A the design of shunts becomes difficult. For ac, current transformers can be used to reduce the current to a level convenient for measurement. At the highest accuracy, current comparators may be used in which flux balance is detected when the magnetizing ampere-turns from two signals are equal and opposite.

Direct-current comparators are available in which dc flux balance is maintained and any unbalance is used to servo a second, or slave, current signal. For the highest accuracy, second-harmonic modulators are used, and for lower precision, Hall effect sensors. Electronically balanced ac and dc current comparators make clip-around ammeters possible, in which an openable magnetic core can be closed around a current-carrying conductor. This allows the meter to be connected into and removed from the circuit without breaking it or interrupting the current. *See* CURRENT COMPARATOR; INSTRUMENT TRANSFORMER.

Very small currents. The obvious method for measuring a very small current is to determine the voltage drop across a large resistor. A sensitive voltage detector having very low offset current is required, for example, an electrometer. Electrometers based on MOSFET (metal-oxide-semiconductor field-effect transistor) devices have overtaken other designs where the very highest resolution is required, as they can have offset current drifts less than 10^{-16} A. In order to provide a low impedance to the measured current, it is preferable to use the electrometer device in an operational-amplifier configuration (see **illus.**). The input becomes a virtual ground, and so stray capacitance across the input connection does not degrade the rate of response of the circuit as seriously as in the simple connection. *See* TRANSISTOR.

The operational-amplifier configuration can be extended to provide a variety of ranges from each high-value resistor. Driving the resistor from the amplifier output through a voltage divider of ratio N gives a similar result to the use of a resistor N times higher in value. This also has the advantage that extremely high-value resistors can be avoided.

Operational amplifiers suitable for this application are readily available, based not only on field-effect transistors but also on varactors. The latter are semiconductor equivalents to the mechanical vibrating capacitor electrometers. *See* AMPLIFIER;



Measurement of very small currents by using an operational amplifier.

ELECTRICAL MEASUREMENTS; ELECTROMETER; OPERATIONAL AMPLIFIER; VARACTOR; VOLTAGE MEASUREMENT. R. B. D. Knight

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Current sources and mirrors

A current source is an electronic circuit that generates a constant direct current which flows into or out of a high-impedance output node. A current mirror is an electronic circuit that generates a current which flows into or out of a high-impedance output node, which is a scaled replica of an input current, flowing into or out of a different node.

Basic circuit configurations. Most specifications of analog integrated circuits depend almost uniquely on the technological parameters of the devices, and on the direct or alternating current that flows through them. The voltage drop over the devices has much less impact on performance, as long as it keeps the devices in the appropriate mode of operation (linear or saturation). High-performance analog integrated-circuit signal processing requires that currents be generated and replicated (mirrored) in an accurate way, independent of supply voltage and of those device parameters that are least predictable (for example, current gain β in a bipolar transistor). Hence, current sources and mirrors occupy a large portion of the total die area of any analog integrated circuit. They are also used, but less often, in discrete analog circuits.

Two basic circuit configurations are shown in **Fig. 1**. The circuit in Fig. 1*a* is suitable for implementation in bipolar integrated-circuit technology, or in complementary metal-oxide-semiconductor (CMOS) integrated-circuit technology if Q_1 and Q_2 are replaced by enhancement metal-oxide-semiconductor field-effect transistors (MOSFETs). Formulas will be derived here only for the bipolar configuration, but can be extended easily to MOS circuit configurations. The collector-base short circuit guarantees that transistor Q_1 will be always in the linear range. If it is assumed, for the moment, that the current gain β and the output impedance r_o of the bipolar transistors are infinite, then the currents are determined by Eqs. (1)–(3). Here, I_{C1} is the collector current of

$$I_{C1} = I_{S1} \cdot \left(\exp \frac{V_{BE1}}{U_T} - 1 \right) = \frac{V_{DD} - V_{SS} - V_{BE1}}{R} \tag{1}$$

$$I_{\text{out}} = I_{S2} \cdot \left(\exp \frac{V_{BE2}}{U_T} - 1 \right) \tag{2}$$

$$V_{BE1} = V_{BE2} = 0.7 \text{ volt} \tag{3}$$

Q_1; V_{BE1} and V_{BE2} are the base-emitter voltages of Q_1 and Q_2; V_{CC} and V_{EE} are the collector and emitter supply voltages; I_{S1} and I_{S2} are the saturation currents of Q_1 and Q_2, which are proportional to the emitter areas A_{E1} and A_{E2}; and U_T is the thermal voltage kT/q, where k is Boltzmann's constant, T is the absolute temperature, and q is the electron charge. (At room temperature, $U_T \simeq 16$ mV.) For two transistors, side by side on the same integrated circuit, the ratio of the two saturation currents I_{S1}/I_{S2} can be realized with an error of less than 1%; Eqs. (1)–(3) can be rewritten as Eq. (4). This equation implies that

$$I_{\text{out}} = \frac{A_{E2}}{A_{E1}} \cdot \frac{V_{DD} - V_{SS} - 0.7 \text{ V}}{R} = \frac{A_{E2}/I_{\text{ref}}}{A_{E1}} \tag{4}$$

the output current will be independent of the technological parameters of the bipolar transistors but dependent on the value of resistor R and on the supply voltages. Even if resistor R or the supply voltages are varied, I_{C1} and I_{C2} always remain proportional: I_{C1} is mirrored in I_{C2}. Equations (1)–(3) demonstrate an important principle in analog integrated-circuit design: replica biasing, in which I_{C1} is forced to produce a base-emitter voltage V_{BE} on transistor Q_1. This voltage creates exactly the right current in Q_2, under the assumption that both halves of the current mirror are exact replicas. The circuit of Fig. 1*a* thus acts as a current source characterized by Eq. (4) if the resistor R is present, and as a current mirror also characterized by Eq. (4) if the resistor is replaced by an input current source of value I_{ref}.

A simple and elegant one-transistor current source can be realized with a junction field-effect transistor (JFET) or depletion MOSFET (Fig. 1*b*). Indeed, the source-to-gate connection guarantees a constant and predictable transistor biasing, since the threshold voltage V_T is negative, as in Eq. (5). The output

$$V_{GS} - V_T = -V_T > 0 \tag{5}$$

current realized in this way is often denoted by I_{DSS}.

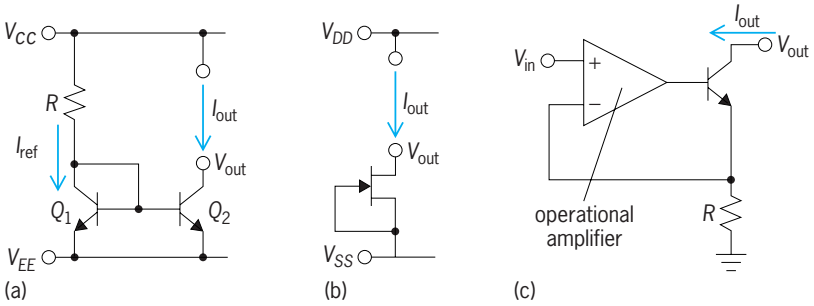

Fig. 1. Basic current-source circuit configurations. (*a*) Circuit suitable for implementation in bipolar or enhancement metal-oxide-semiconductor (MOS) integrated-circuit technology, for example, complementary MOS (CMOS) technology. (*b*) Circuit suitable for implementation in depletion field-effect transistor (FET) integrated-circuit technology, for example, negative MOS (NMOS), bipolar-junction FET (BiFET), or gallium arsenide metal-semiconductor FET (MESFET) technologies. (*c*) More general active circuit (both integrated-circuit and discrete applications).

Finally, a more general circuit configuration (Fig. 1*c*) employs an operational amplifier to create a constant current given by Eq. (6). This circuit

$$I_{\text{out}} = V_{\text{in}}/R \tag{6}$$

has been successfully applied in both integrated and discrete circuits to create constant or voltage-dependent current sources.

Variations. Although the circuit in Fig. 1*a* is the most common circuit configuration, with applications in bipolar and CMOS technology [with bipolars replaced by negative MOS (NMOS) transistors], it has some serious drawbacks that are addressed in the improved circuits in **Fig. 2**.

The output current I_{out} in the circuit of Fig. 1 is affected by the finite current gain β in the bipolar transistors as given by Eq. (7). The circuit in Fig. 2*a*

$$\frac{\Delta I_{\text{out}}}{I_{\text{out}}} = \frac{1}{1 + (2/\beta)} \tag{7}$$

has a considerably lower error due to finite β.

Current mirror ratios I_{S2}/I_{S1} different from unity can be realized by transistor pairs with unequal emitter areas; however, manufacturing tolerances can now lead to significant systematic errors. Integer mirror ratios can be realized without accuracy degradation, if arrays of identical unit transistors are used (3 to 1 in Fig. 2*b*).

The basic current source or current mirror of Fig. 1 has an output impedance equal to the dynamic output resistance r_o of the transistors. Cascodes (Fig. 2*c*) enhance the output resistance of the current source

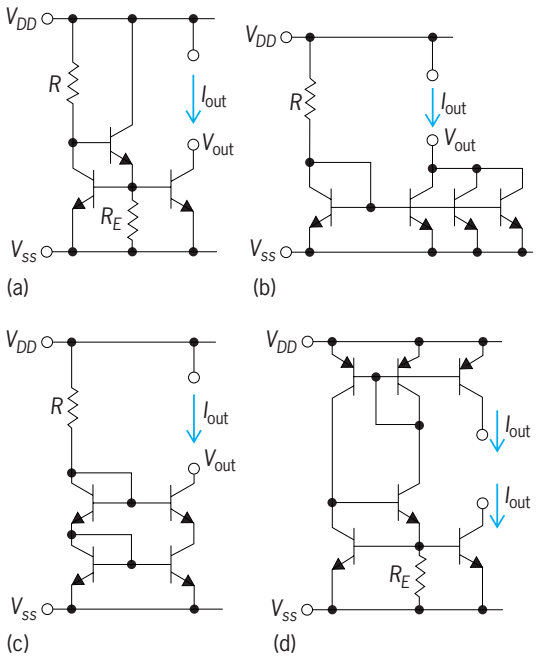

Fig. 2. Improved current-source configurations. (*a*) Circuit for reduction of effect of finite β. (*b*) Replica biasing for integer current mirror ratio. (*c*) Cascoded current mirror for enhanced output impedance. (*d*) Power-supply-independent current source.

according to Eq. (8), where $G_m$is the transconduc-

$$R_{\text{out,source}} = r_o \cdot (G_m \cdot r_o) \tag{8}$$

tance of the cascode transistor. The increase in R_{out} is usually by a factor of about 100.

It is often important that analog circuit specifications be independent of the applied supply voltages. Figure 2*d* shows a simple circuit in which the current is independent of the supply voltages. The realized current I_{out} can be found by solving transcendental equation (9).

$$I_{\text{out}} = V_T \cdot \frac{\log_e(I_{\text{out}}/I_S)}{R_E} \tag{9}$$

This circuit has two operating points: I_{out} given by Eq. (9) or I_{out} equal to zero. To avoid having the circuit lock into the currentless mode during power-up, a startup circuit must be added. *See* INTEGRATED CIRCUITS; TRANSISTOR.

Peter M. VanPeteghem

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Curve fitting

A procedure in which the basic problem is to pass a curve through a set of points, representing experimental data, in such a way that the curve shows as well as possible the relationship between the two quantities plotted. It is always possible to pass some smooth curve through all the points plotted, but since there is assumed to be some experimental error present, such a procedure would ordinarily not be desirable. *See* INTERPOLATION.

The first task in curve fitting is to decide how many degrees of freedom (number of unspecified parameters, or independent variables) should be allowed in fitting the points and what the general nature of the curve should be. Since there is no known way of answering these questions in a completely objective way, curve fitting remains something of an art. It is clear, however, that one must make good use of any background knowledge of the quantities plotted if the above two questions are to be answered. Therefore, if it was known that a discontinuity might occur at some value of the abscissa, one would try to fit the points above and below that value by separate curves.

Against this background knowledge of what the curve should be expected to look like, one may observe the way the points fall on the paper. It may even be advantageous to make a few rough attempts to draw a reasonable curve "through" the points.

A knowledge of the accuracy of the data is needed to help answer the question of the number of degrees of freedom to be permitted in fitting the data. If the data are very accurate, one may use as many degrees of freedom as there are points. The curve can then be made to pass through all the points, and

it serves only the function of interpolation. At the opposite extreme when the data are very rough, one may attempt to fit the data by a straight line representing a linear relation, Eq. (1), between y and x.

$$y = ax + b \tag{1}$$

Using the above information and one's knowledge of the functions that have been found useful in fitting various types of experimental curves, one selects a suitable function and tries to determine the parameters left unspecified. At this point there are certain techniques that have been worked out to choose the optimum value of the parameters.

One of the most general methods used for this purpose is the method of least squares. In this method one chooses the parameters in such a way as to minimize the sum, Eq. (2), where y_i is the ordinate of ith

$$S = \sum_{i=1}^{n} [y_i - f(x_i)]^2 \tag{2}$$

point and $f(x_i)$ the ordinate of the point on the curve having the same abscissa x_i as this point.

Method of least squares. If one attempts to fit the data with a straight line, one lets y equal the values shown in Eq. (3), and Eq. (2) becomes (4).

$$y = f(x) = ax + b \tag{3}$$

$$S = \sum_{i=1}^{n} [y_i - ax_i - b]^2 \tag{4}$$

Since S is to be a minimum,

$$\frac{\partial S}{\partial a} = \frac{\partial S}{\partial b} = 0$$

and therefore Eqs. (5) hold.

$$\begin{aligned} a \sum_i x_i^2 + b \sum_i x_i &= \sum_i x_i y_i \\ a \sum_i x_i + bn &= \sum_i y_i \end{aligned} \tag{5}$$

The solution of these equations is shown as Eqs. (6), where the term Δ is defined in Eq. (7).

$$\begin{aligned} a &= \frac{1}{\Delta}\left(\sum_i x_i \sum_i y_i - n \sum_i x_i y_i\right) \\ b &= \frac{1}{\Delta}\left(\sum_i x_i \sum_i x_i y_i - \sum_i x_i^2 \sum_i y_i\right) \end{aligned} \tag{6}$$

$$\Delta = \left(\sum_i x_i\right)^2 - n \sum_i x_i^2 \tag{7}$$

These values of a and b are those that minimize S and therefore represent the proper choice of the parameters of a straight line according to the criterion of least squares. *See* LEAST-SQUARES METHOD.

Method of averages. This method has less theoretical justification, but it is easier to apply than the method of least squares.

If all the points (x_i, y_i for $i = 1, 2, \ldots, n$) representing the data were to lie on a straight line, one could require that a and b satisfy all n of Eqs. (8). Any two

$$y_i = ax_i + b \qquad i = 1, 2, \ldots, n \tag{8}$$

of these equations could then be used to determine a and b. One could also add these equations together in two groups to obtain Eqs. (9).

$$\begin{aligned} \sum_{i=1}^{k} y_i &= a \sum_{i=1}^{k} x_i + bk \\ \sum_{i=k+1}^{n} y_i &= a \sum_{i=k+1}^{n} x_i + b(n-k) \end{aligned} \tag{9}$$

Suppose now that the points do not lie on a straight line; then Eqs. (8) are no longer valid. One may, however, assume that a best line through the points is given by Eq. (3) with a and b determined by requiring that Eqs. (9) be satisfied. There is some arbitrariness in the choice of k, but it is usually taken to be a whole number near $n/2$.

Consider the data below:

x	y
0.5	0.4
1.0	1.3
1.5	1.6
2.0	2.3
2.5	3.2
3.0	3.6
3.5	4.2
4.0	5.2

Letting $k = 4$, one obtains the following values for Eqs. (9):

$$\begin{aligned} 5a + 4b &= 5.6 \\ 13a + 4b &= 16.2 \end{aligned}$$

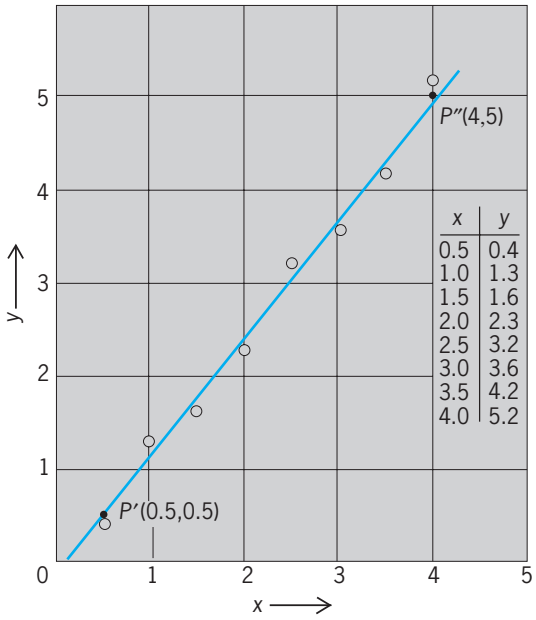
Subtracting, one has $8a = 10.6$ or $a = 1.325$, and therefore $b = -0.256$; thus, the best straight line is given by the equation

$$y = 1.325x - 0.256$$

Graphical method. In this method one first plots the points (x_i,y_i) on coordinate paper using suitable scales along the X and Y axes (see **illus.**). One then takes a transparent straightedge and adjusts it so the points lie as near the edge as possible with about the same numbers above as there are below. The straight line corresponding to the position of the edge is drawn in, and the coordinates of two points (x',y') and (x'',y'') near each end of the line are read off. The straight line drawn is then given by Eq. (3), with a and b determined by the simultaneous equations (10).

$$\begin{aligned} y' &= ax' + b \\ y'' &= ax'' + b \end{aligned} \tag{10}$$

The coordinates of points P' and P'' in the



Graphical method of curve fitting.

illustration are, respectively, (0.5,0.5) and (4,5); therefore, $\alpha = 1.286$ and $b = -0.144$.

Use of nonlinear scales. Suppose one tries to fit the data points (x_i,y_i) with any general functional relation $y = f(x,\alpha,\beta)$ having two undetermined parameters α and β. It often happens that it is possible to introduce two new variables, Eqs. (11), that will satisfy linear Eq. (12). From the given data points (x_i,y_i)

$$X = F(x) \qquad Y = G(y) \tag{11}$$

$$Y = aX + b \tag{12}$$

one can obtain a new set of data points (X_i,Y_i) by the use of Eqs. (11). One may then fit these points by a straight line in the XY plane, given by Eq. (12), using any of the techniques described above. The required equation in x and y is then $G(y) = aF(x) + b$, which will determine the original constants α and β.

For example, suppose one assumes that Eq. (13)

$$y = \beta e^{\alpha x} \tag{13}$$

holds. Then, by taking the natural logarithm of both sides, one has Eq. (14). Thus a linear equation as

$$\log y = ax + \log \beta \tag{14}$$

given in Eq. (12) exists, as shown in Eqs. (15) and (16). A graphical solution involves use of semilog

$$X = x \qquad Y = \log y \tag{15}$$

$$a = \alpha \qquad b = \log \beta \tag{16}$$

paper in which the vertical displacement of points is proportional to $\log y$. On such paper, therefore, Eq. (14), or its equivalent, Eq. (13), plots as a straight line.

Consider next fitting data by Eq. (17), which may

$$y = \beta x^{\alpha} \tag{17}$$

also be written $\log y = \alpha \log x + \log \beta$. Thus again the linear equation labeled (12) connects, as shown in Eqs. (18), and Eqs. (16) again hold for the constants.

$$X = \log x \qquad Y = \log y \tag{18}$$

Logarithmic paper makes it simple to give the data points the rectangular coordinate values X and Y. The best line through these points then determines a and b and, therefore, also α and β.

If the plotting on logarithmic paper seems to lie about a curved line, one may try an equation of the form of Eq. (19), which may be written as Eq. (20).

$$y = \beta x^{\alpha} + \gamma \tag{19}$$

$$\log(y - \gamma) = \alpha \log x + \log \beta \tag{20}$$

One therefore works with a linear relation between $\log x$ and $\log (y - \gamma)$.

To obtain γ one draws a smooth curve through the points (x_i,y_i) and picks out two points (x',y') and (x'',y'') near the ends of this curve. Next one chooses $x''' = {}^1/_2(x'x'')$ and finds the point on the curve (x''',y''') with this value of x. It can be shown quite simply that Eq. (21) holds. The values of a and

$$\gamma = \frac{y'y'' - y'''}{y' + y'' - 2y'''} \tag{21}$$

b and therefore of α and β are found as before.

If one decides to fit the data by a parabolic curve represented by Eq. (22), one may first draw a smooth

$$y = \alpha + \beta x + \gamma x^2 \tag{22}$$

curve through the points and select some point $(x'y')$ on this curve. The parabola will pass through this point if Eq. (23) holds. From Eqs. (22) and (23)

$$y' = \alpha + \beta x' + \gamma x'^2 \tag{23}$$

one has Eq. (24), which shows that $(y - y')/(x - x')$

$$\frac{y - y'}{x - x'} = \beta + \gamma(x + x') \tag{24}$$

is a linear function of x. Therefore, for each data point (x_i,y_i) one may plot a new point (X_i,Y_i), as shown in Eqs. (25). The best straight line through these points

$$X_i = x_i \qquad Y_i = \frac{y_i - y'}{x_i - x'} \tag{25}$$

given by Eq. (12) can be determined as before and this determines β and γ in Eq. (24). The value of α is then determined by use of Eq. (23). *See* COORDINATE SYSTEMS; EXTRAPOLATION; GRAPHIC METHODS.

Kaiser S. Kunz

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Cutaneous sensation

The sensory quality of skin. The skin consists of two main layers: the epidermis, which is the outermost protective layer, and the dermis, which consists of a superficial layer called the papillary dermis and a deeper layer called the reticular dermis (**Fig. 1**). Beneath the dermis is a layer of loose connective tissue, the subcutaneous tissue, which attaches the skin to underlying structures. *See* SKIN.

Sensory Receptors

Sensory receptors in or beneath the skin are peripheral nerve-fiber endings that are differentially sensitive to one or more forms of energy. The sensory endings can be loosely categorized into three morphological groups (**Fig. 2**): endings with expanded tips, such as Merkel's disks found at the base of the epidermis; encapsulated endings, such as Meissner's corpuscles (particularly plentiful in the dermal papillae), and other organs located in the dermis or subcutaneous tissue, such as Ruffini endings, Pacinian corpuscles, Golgi-Mazzoni corpuscles, and Krause's end bulb; and bare nerve endings that are found in all layers of the skin (some of these nerve endings are found near or around the base of hair follicles).

Available evidence indicates that the encapsulated and expanded tip endings and the hair follicle endings are mechanoreceptors and that the specialized end organs serve to influence the dynamic range of mechanical stimuli to which the receptor responds,

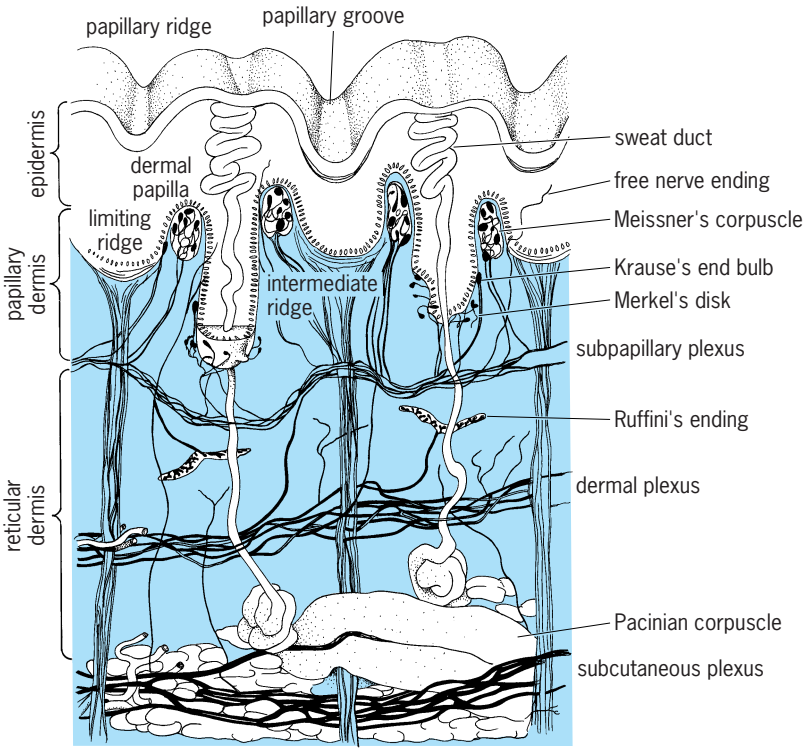

Fig. 1. A section of glabrous (nonhairy) human skin cut vertically through the papillary ridges to show locations of sensory receptors. (*After V. B. Mountcastle, ed., Medical Physiology, vol. 1, 14th ed., C. V. Mosby, 1980*)

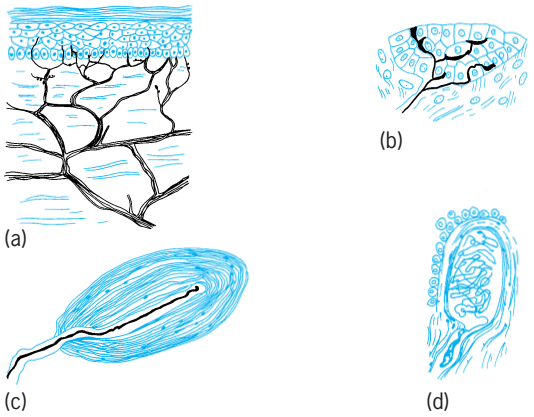

Fig. 2. Types of receptor endings in the skin: (*a*) free nerve endings; (*b*) Merkel's disk; (*c*) Pacinian corpuscle; (*d*) Meissner's encapsulated ending. (*After A. Brodal, Neurological Anatomy in Relation to Chemical Medicine, Oxford University Press, 1969*)

but not the type of energy to which it is most sensitive (for example, mechanical as opposed to thermal). The response specificity of a receptor is determined by some as yet unidentified property of the membrane of the nerve ending. Thus, receptors having bare nerve endings still differ in the specificities of their responses to certain stimuli.

Specificities receptors. There is a remarkable relationship between the response specificities of cutaneous receptors and five primary qualities of cutaneous sensation, the latter commonly described as touch-pressure, cold, warmth, pain, and itch. Each quality is served by a specific set of cutaneous peripheral nerve fibers. More complex sensations must result from an integration within the central nervous system of information from these sets of nerve fibers. For example, tickle sensations may result for the temporal and spatial sequence of stimulating a succession of touch receptors with a slowly moving stimulus lightly applied to the skin. Exploration of the skin surface with a rounded metal point reveals that there exist local sensory spots on the skin, stimulation of which (by any type of energy and by any temporal pattern of stimulation) evokes only one of the five qualities of sensation. Thus may be plotted maps of pressure, warm, cold, pain, or itch spots. *See* PAIN.

There is also a correlation between different qualities of sensation and the sizes (diameters) of the sensory peripheral nerve fibers serving them. The larger-diameter nerve fibers of 6–12 micrometers serve the sense of touch-pressure (and such temporal variations as vibratory sensations or tickle sensations). The sense of coolness and the sense of pricking pain are served by different sets of small myelinated fibers (1–5 μm diameter); sensations of warmth, burning or aching pain, or itch are each mediated by different sets of unmyelinated nerve fibers (0.2–1.5 μm diameter). Since the conduction velocity of nerve impulses along a peripheral nerve is inversely related to the fiber's diameter, there may be a temporal sequence of sensations evoked by an abruptly delivered stimulus. For example, stubbing the toe or pricking

the skin of the foot or hand with a pin may elicit first a sense of touch (mediated by the faster-conducting myelinated fibers), followed almost immediately by a sharp pain of short duration (smaller myelinated fibers) and, after as long as a second or so, a more agonizing pain of longer duration (mediated by the slowly conducting unmyelinated fibers).

Local anesthetic infiltrated around a peripheral nerve first blocks conduction in unmyelinated fibers (with a corresponding loss of the capacity to sense warmth, burning pain, and itch), followed by a block of myelinated fibers in ascending order of their diameters with concomitant losses of the senses of coolness, pricking pain, tickle, and touch-pressure. The exact opposite sequence of sensation loss occurs when a nerve is blocked by asphyxia produced by a pressure cuff around the arm, with the larger-diameter fibers blocked first and the unmyelinated fibers last.

The best evidence for specificity of sensory peripheral nerve fibers and their receptor endings is provided by experiments in humans and animals in which a recording electrode is brought in contact with a portion of the nerve (such as following microdissection of the nerve or by inserting the electrode through the skin and into the nerve), and the electrical signs of impulse discharges of single nerve fibers are recorded in response to cutaneous stimulation. Evidence is that each of the elementary qualities of sensation is served by a different set of peripheral nerve fibers.

Mechanoreceptors. The sense of touch-pressure is mediated by large, myelinated fibers. One type, believed to terminate in Merkel's disks, responds with increasing frequency of discharge to increases in pressure applied to the skin. The frequency of discharge is linearly related to the amount of skin indentation produced by a punctate stimulus and linearly related to numerical estimates, made by human observers, of the perceived magnitude of pressure sensations. Furthermore, the capacity of this mechanoreceptor to respond differentially to gaps of different sizes in a smooth surface applied to the passive skin can account for the human capacity for spatial resolution. The perceptual threshold (minimal separation) for spatially resolving gaps applied to the passive fingertip is a gap of 0.8 mm. Spatial resolution is very good on the fingertip, where the density of mechanoreceptors is high, and relatively poor on the back, where receptor density is sparse.

Other types of mechanoreceptors with large diameter myelinated fibers respond to mechanical movement of the skin and can be considered as velocity detectors. One type ends in Meissner corpuscles and responds best (that is, requires the least amplitude of skin displacement) to mechanical oscillations of 30 Hz. Another type ends in the Pacinian corpuscle and is even more sensitive, but only at higher frequencies (best frequency is about 225 Hz). The combined response properties of receptors with Meissner's or Pacinian corpuscles provide the necessary information required to account, at the level of the receptor, for the capacities of humans to detect and discriminate between mechanical oscillatory stimuli of different frequencies and amplitudes applied to the skin. Such receptors probably have an important role in conveying information about surface texture when the finger is moved back and forth over the surface of an object. *See* MECHANORECEPTORS.

Thermoreceptors. There are two types of thermoreceptors responsive to temperature changes in human skin. Warmth is served by a set of unmyelinated nerve fibers with bare nerve endings. These fibers have an ongoing discharge (in warm skin) that is eliminated by gentle cooling. Their frequencies of discharge increase nearly linearly with increases in stimulus intensity over the range of temperatures described as increasingly warm by human observers. The sense of coolness is served by a set of thinly myelinated nerve fibers with bare nerve endings. These have an ongoing discharge (in cool skin) that is eliminated by gentle warming. They increase their rates of discharge nearly linearly with decreases in temperatures described as increasingly cool by human observers. Neither type of thermoreceptor responds to mechanical stimulation. The response properties of these types of receptors have been shown to provide the information required to account for human capacities to discriminate small differences in the temperatures of stimuli applied to the skin.

Nociceptors. These receptors have bare nerve endings in or beneath the skin and are selectively responsive to noxious stimuli described as painful when applied to human skin. There are different sets of thinly myelinated nerve fibers and unmyelinated fibers whose endings are differentially responsive to only noxious thermal stimuli (hot or cold, or both) or only to noxious mechanical stimuli (such as pricking or pinching the skin). Other types of nociceptors are less specific, and the most common of these is one which responds to any noxious stimulus, whether it be chemical, thermal, or mechanical, and thus is called the polymodal nociceptor. It typically has an unmyelinated nerve fiber, but those with thinly myelinated fibers have been found as well. The minimal intensities of heat required to evoke a response in the typical polymodal nociceptor with an unmyelinated fiber has been shown to elicit minimal sensations of pain in humans. Furthermore, the frequencies of discharge in these receptors in response to heat stimuli increase as a slightly positively accelerating function of increasing temperature over the range of approximately 109–124°F (43–51°C) [higher intensities are not commonly tested], as do numerical estimates by humans of the magnitude of pain evoked by such stimuli. Following injury to the skin, such as that produced by a superficial burn, the threshold responses to mechanical or thermal stimuli of many nociceptors may be greatly lowered. This is believed to account, at least in part, for the increased painfulness of stimuli applied to injured skin. In such a state of heightened sensitivity to pain, termed hyperalgesia, normally innocuous stimulation such as gentle rubbing or warming can evoke a sensation of pain.

Itch. Itch sensations are believed to be mediated by a set of unmyelinated nerve fibers different from those serving pain. Itch can be evoked by mechanical, electrical, and chemical stimuli. Certain chemical stimuli are the most effective in producing itch, and these may be externally applied to the skin or released within the skin under certain conditions. For example, enzymes called proteinases may be found in certain plants and, when delivered to the skin via spicules, are powerful itch-producing agents. A naturally occurring substance is histamine, which is released from mast cells in the skin following local injury or in response to foreign substances that trigger an allergic reaction. Some chemicals produce itch by causing histamine to be released in the skin, while others act independently to evoke itch and can do so in regions of skin previously depleted of histamine.

Robert Lamotte

Touch Sensations

The hands and the oral cavity, including the lips, are the body structures primarily used to tactilely explore small objects in the environment. Touch sensations can easily be aroused from other areas of the body, for example, the stimulation of body hairs, but the resolution of fine detail by these areas is not as great as that of the hands or the oral cavity.

Stimulus characteristics. A mechanical event applied to the skin may be divided into static and dynamic components. During indentation, mechanical energy is expended to move the skin from one position to another. The quantity of energy can be expressed in terms of the applied force per unit dimension, for example, tension (g/mm) or pressure (g/unit area), or in terms of work (force $\times$ indentation depth). Work, compared to pressure or tension, has proved to be a better stimulus measurement to account for variations in touch sensations that are the result of variations in the mechanical properties of the skin. The depth of skin indentation produced by a given force is related to the elastic properties of the skin, which are, in turn, related to age, site of stimulation, and the underlying tissue. If the force is made large and indentation depth controlled, much of the confounding that results from the varying mechanical properties of the skin can be avoided. In this way indentation depth becomes a convenient measurement by which to express the intensities of cutaneous mechanical stimuli.

The velocity, or rate, of indentation has also been shown to be related to threshold tactile sensations. The rate of indentation has little effect on the absolute tactile threshold when the rate of indentation is greater than about 0.3 mm per second. At slower rates of indentation, however, the depth of indentation required to produce a just-detectable tactile sensation increases dramatically. Rate of indentation also has a small but consistent effect on the judged intensity of tactile sensations. Acceleration and deceleration in indentation rate have so far not been implicated as stimulus variables for tactile sensations.

Absolute sensitivity. Sensitivity to tactile stimuli varies from one part of the body to another. The nose, lips, cheek, and forehead are the most sensitive parts, followed closely by the fingertips and palms, but even here there are large differences. Single tactile points on the palmar aspects of the fingers are highly sensitive with a median indentation depth threshold of 0.011 mm, whereas the tactile points in the center of the palm and the lateral aspects of the fingers have a median indentation depth threshold of 0.036 mm.

Intensity of sensations. As the intensity of a touch stimulus (for example, depth of indentation) increases, the intensity of the resulting tactile sensation increases. However, the intensity of the tactile sensation appears to increase in two stages, at least when rates of indentation of 0.4 mm per second or greater are used. Tactile sensations have been judged to be just about the detection level for indentation depths of 0.005 mm and rates of 1 or 10 mm per second. As indentation depth increases, the resulting sensation is judged to increase slowly, but when indentation depths greater than about 0.5 mm are used, the sensation increases considerably more with each stimulus intensity increment. This two-stage response probably results from the cooperation of two different receptor mechanisms to produce the touch sensations.

Spatial discriminations. There are two traditional means of assessing the passive, tactile, spatial, discriminatory capabilites of the skin. These are point localization, the ability to locate the point on the skin that has been touched, and the two-point threshold, the distance separating two pointed stimulators to experience two rather than one point of stimulation. Measurements of the accuracy of point localization show the index finger to have the smallest error (1.4 mm) while the back shows the largest error (12.5 mm). The two-point threshold has been found to be smallest on the middle-finger pad (2.5 mm, the approximate separation of braille dots). The next smallest two-point threshold was on the index-finger pad (3.0 mm), and largest on the calf of the leg (47.0 mm). This means that the finger pads can detect much smaller (finer) features of an object than can other skin areas.

Temporal discriminations. When rapidly vibrating mechanical stimuli are applied to the skin, another stimulus dimension is added—the frequency of repetition. When applied to the index-finger pad, the resulting sensation, for repetition frequencies up to about 40 Hz, may best be described as a flutter. At higher frequencies the flutter changes to a buzzing sensation.

The lower limit to which the skin is sensitive is defined as that frequency at which the individual pulses fuse to form a single continuous sensation. This occurs, as it does in audition, at about 15 Hz. The upper limit is defined as that frequency at which the buzzing sensation disappears and the sensation becomes that of steady skin indentation, about 640 Hz.

Thresholds of vibratory sensations vary with the frequency of the vibration. There is little change in sensitivity for frequencies up to about 40 Hz. As the

frequency of vibration increases beyond 40 Hz, the threshold decreases; that is, the receptors are more sensitive to the higher frequencies, until they reach their peak sensitivity at about 250 Hz, beyond which their sensitivity decreases. Sensitivity to increased frequency once again shows two stages. Two separate receptor processes, each with different rules of operation, combine to form the range of response.

For many years scientists and engineers have sought ways in which the skin might be used as a substitute means of communication for the blind or deaf. There are difficulties, such as cost, ease of learning, and the cosmetic appearance, in the development of a suitable device. In addition, the skin presents its own difficulties in terms of the amount of information that it can convey to the brain. Three information channels are possible: body site of stimulation, frequency, and intensity of the vibratory stimuli. The reliability of point localization is poor except on the fingertips and face, and varies widely with body site. Frequency requires discrimination by the skin—vibrations of 40 and 44 or 200 and 220 Hz applied to the finger pad can be discriminated about 75% of the time, but like point localization, this varies with the site of stimulation. The range of useful frequencies is also severely limited—up to about 600 Hz. Discrimination between intensities of vibration not only varies with the site of stimulation but also varies over a wide range with the frequency of the vibration.

Temperospatial sensations. When a light tactile stimulus, such as the tip of a feather, is drawn across the skin so as to stimulate successive touch points in temporal succession, a tickle sensation is often experienced. This is a different sensation than that of itch, although the response elicited may be the same—rubbing or scratching. Little is known of the necessary and sufficient stimulus conditions to arouse a tickle.

Active touch. The limits of fine discrimination (two-point discrimination) discussed above involved the finger pads resting passively on the feature to be detected. Much finer features (texture) can be detected, however, if the finger pads are actively moved over the surface. Scanning an object with the finger increases the information flow in a tactile sensation. It is not the movement of the finger per se over the object, for just as fine a discrimination can be made if the object is moved under the stationary finger. Perhaps small vibrations, shearing forces on the skin, the temporal pattern of tactile receptor stimulation, or some similar dimensions have been added by the movement.

Another, more complex aspect of active touch is object recognition, or stereognosis. Here, not only are the fine features of the object distinguished, but so is the object's overall form. Kinesthesis, the muscle and joint sense, is used.

Thermal Sensations

The thermal sense is unique among the senses in that an absence of sensation occurs at about 86–97°F (30–36°C). This represents a considerable amount of thermal energy, but this is the temperature domain in which homeothermic systems operate. Changes in temperature imposed from outside change the thermal environment of the (warm and cold) receptors, presumably producing changes in their metabolism and hence their response rates, without an exchange of energy between the stimulus temperature (change) and the receptor. In other sensory systems the absence of the appropriate energy form leads to the absence of sensation. In these systems stimulation is the result of an exchange of energy between the stimulus and the receptor.

The thermal sense is unique in another way. Its primary function—an important input to the body temperature–regulating system—does not reach awareness. Only relatively innocuous sensations of warm and cool are perceived, and only when the skin temperature drops below about 64°F or 18°C (cold-pain) or rises above 113–115°F or 45–46°C (heat-pain) does the sensation become noxious. This represents engagement of a different system—the nocifensor system—and is not properly a part of the thermal sensing system.

Stimulus characteristics. There are six principal stimulus variables that alter the responsiveness of the human temperature-sensing system: (1) the temperature to which the skin has been adapted; (2) the direction of the temperature change; (3) the intensity of the temperature change; (4) the rate of the temperature change; (5) the area of the skin stimulated; and (6) the site of application of the thermal stimuli.

The band of skin temperatures between about 86 and 97°F (30 and 36°C) is called the zone of physiological zero (**Fig. 3**). When skin temperatures do not exceed the limits of this band, given time, thermal sensations adapt completely; that is, they disappear. At lower skin temperatures a cool sensation persists no matter how long that temperature is maintained—the stimulus adapts but the sensation never disappears completely. With further reductions in skin temperature the sensation would be described as cold, and finally at about 64°F (18°C) is replaced by a dull, aching sensation—cold-pain. When the skin temperature is raised above the zone of physiological zero, much the same series of sensations occur except that their qualities are warm to hot, and finally at about 113°F (45°C) the hot is replaced by a stinging sensation—heat-pain.

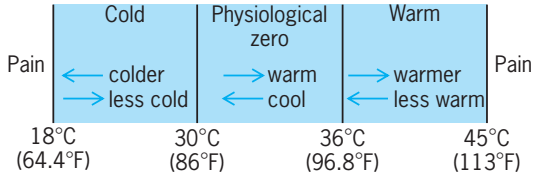

Fig. 3. The relationship between skin temperature, physiological zero, and the direction of temperature change as they affect thermal sensations. (*From D. R. Kenshalo, Sr., Biophysics and psychophysics of feeling, in E. C. Carterette and M. P. Friedman, eds., Handbook of Perception, vol. 6B, chap. 2, Academic Press, 1978*)

The direction of a temperature change, when skin temperature is within the zone of physiological zero, produces thermal sensations consistent with the direction of the change—an increase produces a warm sensation and a decrease produces a cool sensation. A small increase in temperature from a skin temperature below physiological zero produces a decrease in the persisting cool sensation, not a warm sensation. Similarly, a small decrease in temperature from a skin temperature above physiological zero produces a decrease in the persisting warm sensation, not a cool sensation.

The intensity of a temperature change, when skin temperature is within the zone of physiological zero, produces sensations of proportional intensity. When skin temperature is below the zone of physiological zero, sensitivity to decreases in temperature remains unchanged from that in the zone of physiological zero. Sensitivity to temperature increases is reduced, and the farther the skin temperature is below the lower limit of physiological zero, the less sensitive it becomes to temperature increases. A similar circumstance occurs for temperature decreases when the skin temperature is above the upper limit of physiological zero.

When the rate of a temperature change is less than about 0.05°F (0.03°C) per second, sensitivity to that change decreases. At the extreme, skin temperature can be changed from one to the other limit of physiological zero without producing a thermal sensation if done sufficiently slowly, about 0.013°F (0.007°C) per second. Sensitivity to temperature changes does not increase at rates of change faster than 0.05°F (0.03°C) per second, at least up to rates of 2.7°F (1.5°C) per second.

The size of the area of the skin stimulated by temperature changes markedly influences sensitivity. For temperature increases, areal summation is almost complete for areas up to about 30 in.^{2} (200 cm^2). Thus, there is almost complete reciprocity between the area stimulated and the intensity of the stimulus, so that if area is doubled the stimulus intensity required to reach detectability will be halved. This means that the whole palm of the hand is a much more sensitive heat detector than just the fingertip. Areal summation for temperature decreases is somewhat less than that for temperature increases but is still considerable. Such generous summation is not without a price. The thermal sensing system is almost devoid of any ability for spatial localization. Subjects are able to judge only with 80% accuracy whether radiant heat stimuli were delivered to the back or abdomen. This is easy to understand when one considers that the primary function of the cutaneous temperature-sensing system is to inform the central temperature-regulating system of the heat load experienced at the body surface rather than to specify the spatial location of warm and cold objects.

Various parts of the body differ in their sensitivity to temperature increases and decreases. For small temperature increases the forehead is the most sensitive, followed by the chest, forearm, calf, thigh, and abdomen. For small temperature decreases the body trunk is the most sensitive, followed by the limbs and finally the face. With relatively large changes in skin temperature (both increases and decreases) such regional differences in sensitivity are markedly reduced.

Paradoxical thermal sensations. Some cold points, when touched by a hot thermal probe (113°F or 45°C), will yield a cold sensation. The phenomenon of a hot probe giving rise to a cold sensation is called paradoxical cold. Paradoxical cold sensations do not occur, however, when a hot stimulus is applied over a large area. The necessary and sufficient conditions for paradoxical warmth (arousal of a warm sensation by application of a cold probe to a warm point) have not been established.

Heat sensations. The sensation of heat has been held to be uniquely different from either warm or heat-pain sensations. Its threshold ranges from 104 to 115°F (40 to 46°C) with a mean at about 108–109°F (42–43°C).

Two theories dealing with the mechanisms of heat sensations have been advanced. The first maintains that heat sensations are synthesized from the simultaneous stimulation of warm and paradoxically stimulated cold receptors. The second theory maintains that heat sensations arise from their own specifically sensitive receptors. Neither theory has been satisfactorily verified or refuted. *See* NERVOUS SYSTEM (VERTEBRATE); SENSATION; SOMESTHESIS.

Dan R. Kenshalo, Sr.

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Cyanamide

A term used to refer to the free acid, H_2NCN, or commonly to the calcium salt of this acid, $CaCN_2$, which is properly called calcium cyanamide. Calcium cyanamide is manufactured by the cyanamide process, in which nitrogen gas is passed through finely divided calcium carbide at a temperature of 1000°C (1830°F).

Most plants are equipped to produce the calcium cyanamide from the basic raw materials of air, limestone, and carbon according to reactions (1)–(3).

$$CaCO_3 \xrightarrow{\Delta} CaO + CO_2 \quad (1)$$

$$CaO + 3C \longrightarrow CaC_2 + CO \quad (2)$$

$$CaC_2 + N_2 \rightleftharpoons CaCN_2 + C \quad (3)$$

Most calcium cyanamide is used in agriculture as a fertilizer; some is used as a weed killer, as a pesticide, and as a cotton defoliant.

Cyanamide, H_2NCN, is prepared from calcium cyanamide by treating the salt with acid. It is used in the manufacture of dicyandiamide and thiourea. *See* CARBON; NITROGEN.

E. Eugene Weaver

Cyanate

A compound containing the —OCN group, typically a salt or ester of cyanic acid (HOCN). The cyanate ion is ambidentate, that is, it has two reactive sites, because it can bind through the oxygen (O) or the nitrogen (N). Cyanate is commonly N-bonded with most nonmetallic elements, presumably because of the small charge density on the oxygen. Cyanic acid has the structure H—O═C≡N, but it may exist in an isomeric form known as isocyanic acid, H—N═C═O. The cyanates are isomeric with fulminates, where the carbon and the nitrogen are transposed (—ONC).

Cyanic acid is a volatile liquid, and it polymerizes upon standing to form cyamelide acid and cyanuric acid. In water cyanic acid undergoes hydrolysis to ammonia and carbon dioxide. However, dilute solutions in ice water may be kept for several hours. Both cyanic acid and isocyanic acid are prepared from cyanuric acid. The linear —OCN ion may be prepared by mild oxidation of aqueous cyanide ion (CN^-) using lead oxide (PbO) and potassium cyanide (KCN).

The primary use for cyanates is in the synthesis of a number of organic compounds such as unsubstituted carbamates (R—NH═C═OOR′). These compounds are used as derivatives of alchols. Other important reactions include addition to amines to give substituted ureas. *See* CARBON; CYANIDE; NITROGEN; THIOCYANATE.

Thomas J. Meade

Cyanide

A compound containing the —CN group, for example, potassium cyanide, KCN; calcium cyanide, $Ca(CN)_2$; and hydrocyanic (or prussic) acid, HCN. Chemically, the simple inorganic cyanides resemble chlorides in many ways. Organic compounds containing this group are called nitriles. Acrylonitrile, CH_2CHCN, is an important starting material in the manufacture of fabrics, plastics, and synthetic rubber.

HCN is a weak acid, having an ionization constant of 1.3×10^{-9} at 18°C (64°F). In the pure state, it is a highly volatile liquid, boiling at 26°C (79°F).

HCN and the cyanides are highly toxic to animals, a lethal dose to humans usually considered to be 100–200 mg (0.0035–0.007 oz). Death has been attributed, however, to as low a dose as 0.57 mg of HCN per kilogram (9×10^{-6} oz per pound) of body weight. As poisons, they are rapid in their action. When these compounds are inhaled or ingested, their physiological effect involves an inactivation of the cytochrome respiratory enzymes, preventing tissue utilization of the oxygen carried by the blood.

The cyanide ion forms a variety of coordination complexes with transition-metal ions. Representative complexes are those of gold $[Au(CN)_2]^-$, silver $[Ag(CN)_2]^-$, and iron $[Fe(CN)_6]^{4-}$. This property is responsible for several of the commercial uses of cyanides. In the cyanide process, the most widely used method for extracting gold and silver from the ores, the finely divided ore is contacted with a dilute solution of sodium or potassium cyanide. The metal is solubilized as the complex ion, and, following extraction, the pure metal is recovered by reduction with zinc dust. In silver-plating, a smooth adherent deposit is obtained on a metal cathode when electrolysis is carried out in the presence of an excess of cyanide ion. *See* ELECTROPLATING OF METALS; GOLD METALLURGY; SILVER METALLURGY.

$Ca(CN)_2$ is extensively used in pest control and as a fumigant in the storage of grain. In finely divided form, it reacts slowly with the moisture in the air to liberate HCN.

In case hardening of metals, an iron or steel article is immersed in a bath of molten sodium or potassium cyanide containing sodium chloride or carbonate. At temperatures of 750°C (1380°F) and above, the cyanide decomposes at the surface, forming a deposit of carbon which combines with and penetrates the metal. The nitrogen also contributes to the increased hardness at the surface by forming nitrides with iron and alloying metals. *See* SURFACE HARDENING OF STEEL.

NaCN and KCN are produced commercially by neutralization of hydrogen cyanide with sodium or potassium hydroxide. The neutralized solution must then be evaporated to dryness in a strictly controlled manner to avoid undue losses. $Ca(CN)_2$ is produced primarily from calcium cyanamide by reaction with carbon in the presence of sodium chloride at 1000°C (1830°F). *See* COORDINATION CHEMISTRY; FERRICYANIDE; FERROCYANIDE; TOXICOLOGY.

Francis J. Johnston

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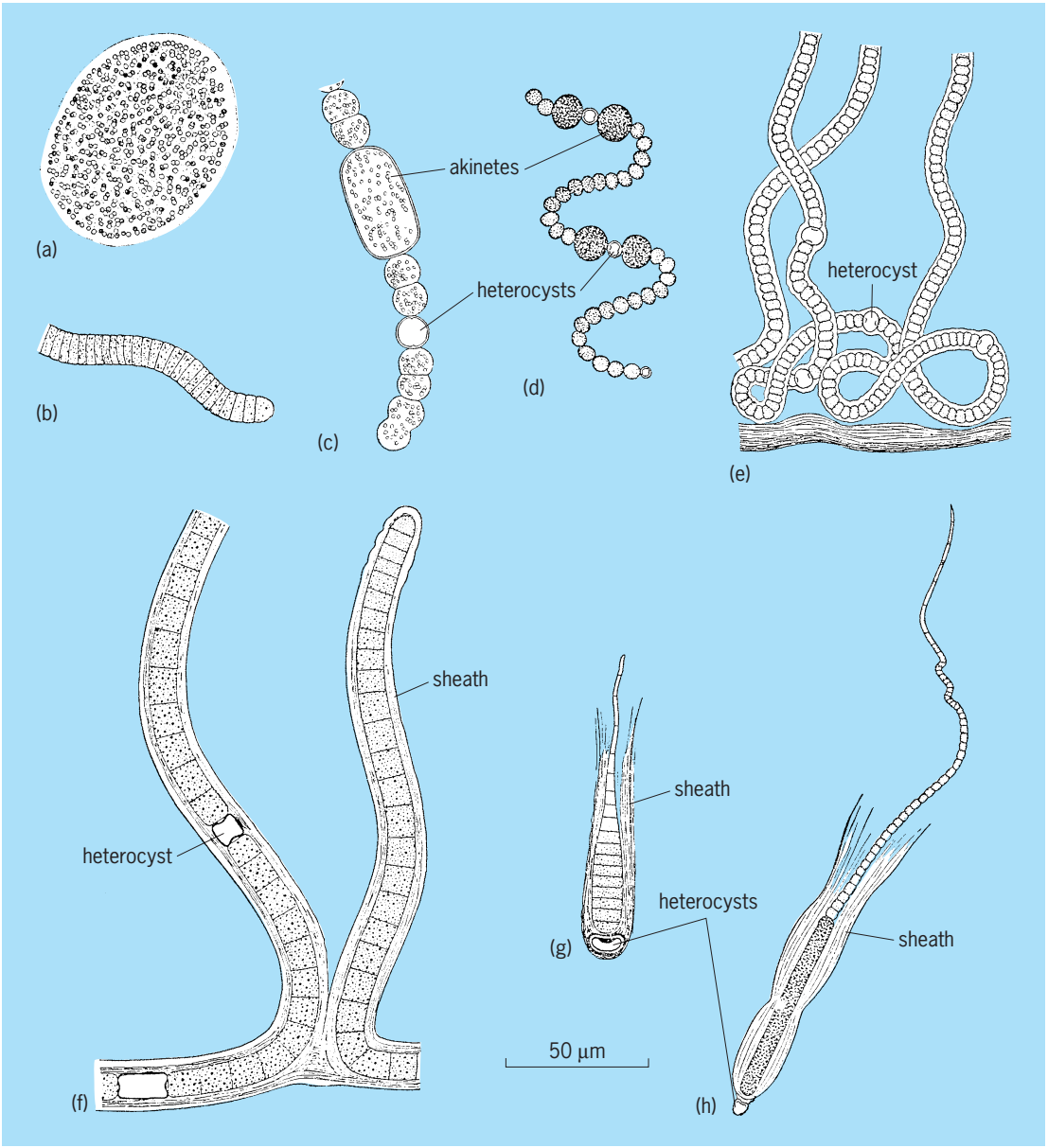
Cyanobacteria

A large and heterogeneous group of photosynthetic microorganisms, formerly referred to as blue-green algae. They had been classified with the algae because their mechanism of photosynthesis is similar to that of algal and plant chloroplasts; however, the cells are prokaryotic, whereas the cells of algae and plants are eukaryotic. The name cyanobacteria is now used to emphasize the similarity in cell structure to other prokaryotic organisms. *See* ALGAE; CELL PLASTIDS.

General characteristics. All cyanobacteria can grow with light as an energy source, generating energy and reducing power through oxygen-evolving photosynthesis; carbon dioxide (CO_2) is fixed into organic compounds via the Calvin cycle, the same mechanism used in green plants. Thus, all species will grow in the absence of organic nutrients. However, some species will assimilate organic compounds into cell material if light is available, and a few isolates are capable of growth in the dark by using organic compounds as carbon and energy sources. Some cyanobacteria can shift to a different mode of photosynthesis, in which hydrogen sulfide rather than water serves as the electron donor. Molecular



Morphological diversity of cyanobacteria. (*a*) *Microcystis* sp.; (*b*) *Oscillatoria* sp.; (*c, d*) *Anabaena* sp.; (*e*) *Nostoc* sp. with gelatinous capsule; (*f*) *Scytonema* sp.; (*g*) *Calothrix* sp.; (*h*) *Gleotrichia* sp. (*After L. H. Tiffany and M. E. Britton, The Algae of Illinois, University of Chicago Press, 1952*)

oxygen is not evolved during this process, which is similar to that in purple and green photosynthetic sulfur bacteria. *See* PHOTOSYNTHESIS.

The photosynthetic pigments of cyanobacteria include chlorophyll *a* (also found in algae and plants) and phycobiliproteins. One of these latter pigments, *c*-phycocyanin, is blue, and led to the name cyanobacteria (Greek cyano-, blue), and in combination with the green chlorophyll pigment gives cells a blue-green coloration. However, other species also contain a red phycobiliprotein, phycoerythrin, and these are red or brown in color. The coloration of the organisms can also be affected by the presence of orange or yellow carotenoid pigments. *See* CAROTENOID; CHLOROPHYLL.

Form and structure. Cyanobacteria have limited metabolic versatility in comparison with other bacteria, but they are extremely diverse morphologically. Species may be unicellular or filamentous (see **illus.**). Both types may aggregate to form macroscopically visible colonies. The cells range in size from those typical of bacteria (0.5–1 micrometer in diameter) to 60 μm in *Oscillatoria princeps*. The latter is among the largest-sized cells known in prokaryotes.

When examined by electron microscopy, the cells of cyanobacteria appear similar to those of gram-negative bacteria. The cell wall contains an outer lipid membrane and a layer of peptidoglycan. Many species produce extracellular mucilage or sheaths

that promote the aggregation of cells or filaments into colonies.

The photosynthetic machinery is located on internal membrane foldings called thylakoids. Chlorophyll *a* and the electron transport proteins necessary for photosynthesis are located in these lipid membranes, whereas the water-soluble phycobiliprotein pigments are arranged in particles called phycobilisomes which are attached to the lipid membrane.

Several other types of intracellular structures are found in some cyanobacteria. Gas vesicles are rigid shells of protein constructed in a way to exclude water from the interior. The hollow structures, which have a diameter of about 70 nanometers and lengths up to 1000 nm, have densities much less than that of water and thus may confer buoyancy on the organisms. They are often found in cyanobacteria that grow in the open waters of lakes. Polyhedral bodies, also known as carboxysomes, contain large amounts of ribulose bisphosphate carboxylase, the key enzyme of CO_2 fixation via the Calvin cycle. Several types of storage granules may be found; glycogen and polyphosphate granules are also found in other organisms, but cyanophycin granules, which consist of polymers of the amino acids aspartic acid and arginine, are unique to cyanobacteria. These storage granules serve as reserves of carbon and energy, phosphorus, and nitrogen, respectively. They are synthesized when the particular nutrient is present in excess in the environment, and the reserves are utilized when external sources of the nutrient are unavailable.

Specialized cells. Some filamentous cyanobacteria can undergo a differentiation process in which some of the normal vegetative cells become heterocysts. The process occurs when combined nitrogen (ammonia and nitrate) is depleted in the environment. Heterocysts contain the enzyme nitrogenase, which catalyzes the conversion of nitrogen gas to ammonia. Nitrogen fixed in the heterocysts is translocated to the vegetative cells in the filament, to allow their continued growth in the absence of external ammonia or nitrate. Other changes occur during the differentiation into heterocysts. The cell wall becomes thicker, and the heterocyst develops intracellular connections with adjacent vegetative cells. In addition, the heterocysts lose the portion of the photosynthetic apparatus responsible for oxygen evolution. These changes are necessary for nitrogen fixation to occur, because nitrogenase is inactivated by oxygen. Thus, the heterocyst becomes an anaerobic site for nitrogen fixation, and depends upon translocation of organic compounds from vegetative cells to provide the reducing power to drive nitrogen fixation. *See* NITROGEN FIXATION.

In some filamentous species, resting spores called akinetes are formed. These cell forms are more resistant to periods of drying, freezing, or darkness than are the normal vegetative cells.

Occurrence. Cyanobacteria can be found in a wide variety of freshwater, marine, and soil environments. They are more tolerant of environmental extremes than are eukaryotic algae. For example, they are the dominant oxygenic phototrophs in hot springs (at temperatures up to 72°C or 176°F) and in hypersaline habitats such as may occur in marine intertidal zones. In these areas, they are often components of microbial mats, layers of microbial protoplasm up to several centimeters in thickness. In addition to cyanobacteria, other phototrophic bacteria and heterotrophic bacteria proliferate in these mats.

Cyanobacteria are often the dominant members of the phytoplankton in fresh-water lakes that have been enriched with inorganic nutrients such as phosphate. Up until the 1970s, cyanobacteria were not thought to be important members of the oceanic plankton, but it is now known that high population densities of small, single-celled cyanobacteria occur in the oceans, and that these are responsible for 30–50% of the CO_2 fixed into organic matter in these environments. *See* PHYTOPLANKTON.

Some cyanobacteria form symbiotic associations with eukaryotic photosynthetic organisms. These associations are based on the capacity of heterocystous cyanobacteria to fix nitrogen. *Nostoc* species reside in specialized cavities on the undersurface of liverworts, such as *Anthoceros*. In this environment, 30–40% of the *Nostoc* cells become heterocysts (a frequency tenfold higher than in free-living species), and 90% of the nitrogen fixed in the heterocysts is assimilated by the liverwort. Associations also occur with the water fern *Azolla*, some cycads, and the angiosperm *Gunnera*.

Lichens are symbiotic associations between a fungus and a photosynthetic microorganism. About 8% of the lichens involve a cyanobacterium, which can provide both fixed nitrogen and fixed carbon to the fungal partner. *See* LICHENS.

Dispersal. Cyanobacteria in open waters or in microbial mats can adjust their vertical position in the environment. Planktonic cyanobacteria do so by regulating the buoyancy conferred by gas vesicles. These cells can lose buoyancy by accumulating dense polymers such as glycogen. Filamentous cyanobacteria in microbial mats are often capable of gliding motility, a process in which cells lacking flagella can move in contact with surfaces. In both instances, the organisms adjust their vertical positions to avoid high, damaging light irradiances. However, their behavior is also modulated by the presence of inorganic nutrients or toxic substances such as hydrogen sulfide.

Evolution. Cyanobacteria are thought to be the first oxygen-evolving photosynthetic organisms to develop on the Earth, and hence responsible for the conversion of the Earth's atmosphere from anaerobic to aerobic about 2 billion years ago. This development permitted the evolution of aerobic bacteria, plants, and animals. *See* BACTERIA; CYANOPHYCEAE; PREBIOTIC ORGANIC SYNTHESIS. Allan Konopka

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Photosynthetic Picoplankton, 1987; B. A. Whitton and N. G. Carr (eds.), *The Biology of Cyanobacteria*, 1982.

Cyanocarbon

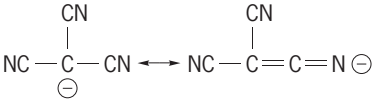
A derivative of hydrocarbon in which all of the hydrogen atoms are replaced by the —C≡N group. Only two cyanocarbons, dicyanoacetylene and dicyanobutadiyne, were known before 1958. Since then, tetracyanomethane, hexacyanoethane, tetracyanoethylene, hexacyanobutadiene, and hexacyanobenzene were synthesized.

The term cyanocarbon has been applied to compounds which do not strictly follow the above definition: tetracyanoethylene oxide, tetracyanothiophene, tetracyanofuran, tetracyanopyrrole, tetracyanobenzoquinone, tetracyanoquinodimethane, tetracyanodithiin, pentacyanopyridine, diazomalononitrile, and diazotetracyanocyclopentadiene

Tetracyanoethylene, the simplest olefinic cyanocarbon, is a colorless, thermally stable solid, having a melting range of 198–200°C (388–392°F). It is a strong π-acid, readily forming stable complexes with most aromatic systems. These complexes absorb radiation in or near the visible spectrum.

Charge-transfer salts derived from tetracyanoquinodimethane and tetrathiafulvalene are highly conducting and have attracted much interest as possible replacements for metal conductors.

Cyanocarbon acids are among the strongest protonic organic acids known and are usually isolated only as the cyanocarbon anion salt. These salts are usually colored and stabilized by resonance of the type indicated in the reaction below.



Twenty-five similar forms can be written for the 2-dicyanomethylene-1,1,3,3-tetracyanopropanediide ion, $(NC)_2C{=}C[C(CN)_2]$, which accounts for the similarity between the acid strength of the free acid and that of sulfuric acid.

The reactivity of cyanocarbons is illustrated by the facile replacement of one of the cyano groups in tetracyanoethylene by nucleophilic attack under very mild conditions and by its addition to dienes. Examples of these reactions are shown in the **illustration**. *See* ORGANIC CHEMISTRY.

Owen W. Webster

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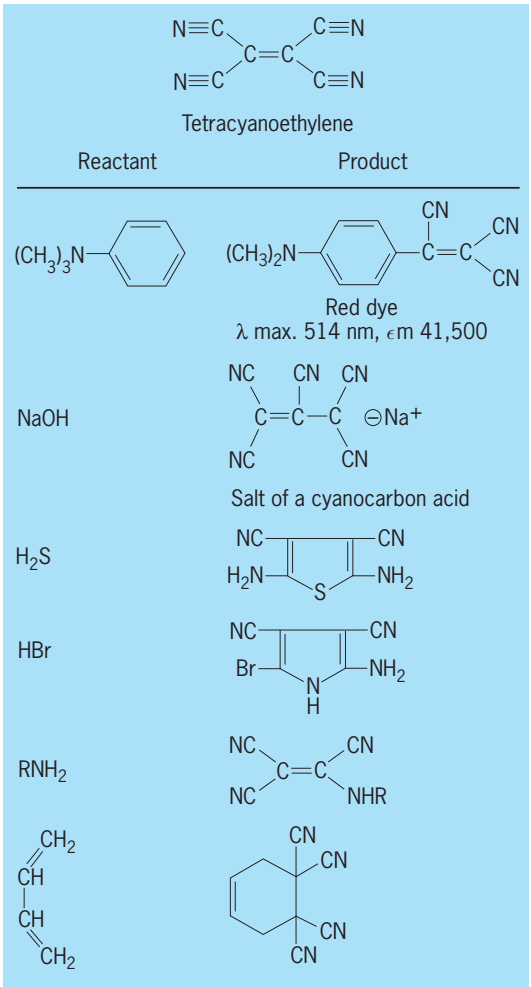


Reactions of tetracyanoethylene.

Cyanogen

A colorless, highly toxic gas having the molecular formula C_2N_2. Cyanogen belongs to a class of compounds known as pseudohalogens, because of the similarity of their chemical behavior to that of the halogens. Liquid cyanogen boils at −21.17°C (−6.11°F) and freezes at −27.9°C (18.2°F) at 1 atm (10^5 Pa). The density of the liquid is 0.954 g/ml at the boiling point.

Cyanogen may be prepared by prolonged heating of mercuric cyanide at 400°C (750°F), reaction (1), or by allowing a solution of copper sulfate to

$$Hg(CN)_2 \rightarrow Hg + C_2N_2 \qquad (1)$$

flow slowly into a solution of potassium cyanide, reaction (2). The copper(II) cyanide is unstable, de-

$$2KCN + CuSO_4 \rightarrow Cu(CN)_2 + K_2SO_4 \qquad (2)$$

composing to give cyanogen and copper(I) cyanide, reaction (3).

$$2Cu(CN)_2 \rightarrow 2CuCN + C_2N_2 \qquad (3)$$

Cyanogen is also formed by the reaction in the gas phase of hydrogen cyanide and chlorine at

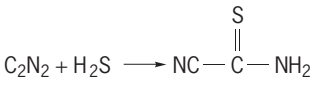
400–700°C (750–1300°F) in the presence of a catalytic agent such as activated carbon, reaction (4).

$$2HCN + Cl_2 \rightarrow C_2N_2 + 2HCl \quad (4)$$

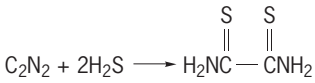
Cyanogen chloride, ClCN, is an intermediate in this reaction. When heated to 400°C (750°F), cyanogen gas polymerizes to a white solid, paracyanogen $(CN)_x$.

Cyanogen reacts with hydrogen at elevated temperatures in a manner analogous to the halogens, forming hydrogen cyanide, reaction (5). With hydrogen sulfide, H_2S, cyanogen forms thiocyanoformamide or dithiooxamide. These reactions are shown by (6) and (7).

$$C_2N_2 + H_2S \rightleftharpoons 2HCN \quad (5)$$



$$C_2N_2 + H_2S \longrightarrow NC-\overset{\overset{S}{\|}}{C}-NH_2 \quad (6)$$



$$C_2N_2 + 2H_2S \longrightarrow H_2NC-\overset{\overset{S}{\|}}{\underset{\underset{S}{\|}}{C}}-CNH_2 \quad (7)$$

Cyanogen burns in oxygen, producing one of the hottest flames known from a chemical reaction. It is considered to be a promising component of high-energy fuels.

Structurally, cyanogen is written N≡C—C≡N. *See* CYANIDE.

Francis J. Johnston

Bibliography. R. J. Fressenden and J. S. Fressenden, *Organic Chemistry*, 5th ed., 1994; Z. Rappaport (ed.), *The Chemistry of the Cyano Group*, 1971.

Cyanophyceae

A class of prokaryotic organisms coextensive with the division Cyanophycota of the kingdom Monera. Because these organisms have chlorophyll *a* and carry out oxygen-evolving photosynthesis, they have traditionally been aligned with algae and, with regard for their characteristic color, called blue-green algae. Microbiologists have emphasized the prokaryotic structure of these organisms and aligned them with bacteria, as the Cyanobacteria. Other names applied to these organisms include Cyanophyta at the level of division and Myxophyceae or Schizophyceae at the level of class. Blue-greens range in form from unicells 1–2 micrometers in diameter to filaments 10 cm (4 in.) long. *See* CYANOBACTERIA.

Classification. The taxonomy of the blue-greens is in an especially unsteady state, polarized by two viewpoints. At one pole (Geitlerian system) nearly every morphological variant encountered in nature is accorded taxonomic recognition. At the other pole (Drouetian system) great morphological plasticity is attributed to each species. Geitler recognized 4 orders, 22 families, 140 genera, and more than 1200 species, while Drouet recognized 2 orders, 6 families, 24 genera, and 61 species. Most specialists of Cyanophyceae steer a middle course. A further major consideration has recently been introduced by microbiologists, who have proposed their own system of classification based solely on characters revealed in pure cultures.

Unicellular forms, which may aggregate in colonies or loosely constructed filaments and which reproduce by binary fission or spores, constitute the order Chroococcales, from which two additional orders, Chamaesiphonales and Pleurocapsales, are sometimes segregated. Filamentous forms, which reproduce by hormogonia, constitute the order Nostocales (= Hormogonales or Oscillatoriales). The Nostocales may be restricted to unbranched or falsely branched forms (in which the ends of a trichome adjacent to a rupture grow out as a pair of branches), while those forms in which cells divide in more than one plane (true branching) constitute the order Stigonematales.

Habitat. Geographically and ecologically, blue-green algae are nearly as ubiquitous as bacteria. They are especially abundant in the plankton of neutral or alkaline eutrophic fresh waters and tropical seas, often forming blooms. Habitats for benthic forms include hot springs, snow and ice, soil, rocks, tree trunks, and buildings. Cyanophyceae live symbiotically with a large variety of animals and plants, including sponges, diatoms, liverworts, cycads, the aquatic fern *Azolla*, and the flowering plant *Gunnera*. They constitute the phycobiont of many lichens.

In addition to contributing to food chains, blue-green algae play specific beneficial roles. Nitrogen-fixing forms greatly enrich rice paddies. *Spirulina*, a traditional food in parts of Mexico and central Africa, is grown commercially and marketed widely as a high-protein dietary supplement. On the other hand, blue-green algae are often a nuisance. They clog filters, impart undesirable tastes and odors to domestic water supplies, and make unusable or at least unattractive many swimming pools, aquariums, and fountains. Cyanophycean blooms are often toxic to fish, birds, and livestock.

Blue-green algae were pioneers on Earth and are known from rocks at least as old as 2.3 billion years. They are believed to have been responsible for the accumulation of oxygen in the primeval atmosphere and to have been involved in the formation of laminated reeflike structures called stromatolites. To judge from similar modern structures (as at Shark Bay, Western Australia), stromatolites were produced by algal mats that entrapped detrital sediments and sometimes deposited calcium carbonate. *See* ALGAE; STROMATOLITE.

Paul C. Silva; Richard L. Moe

Bibliography. H. C. Bold and M. J. Wynne, *Introduction to the Algae: Structure and Reproduction*, 1985; F. Drouet, Revision of the Stigonemataceae with a summary of the classification of the blue-green algae, *Beih. Nova Hedwigia*, Heft 66, 1981; S. P. Parker (ed.), *Synopsis and Classification of Living Organisms*, 2 vols., 1982.

Cybernetics

The study of communication and control within and between humans, machines, organizations, and society. This is a modern definition of the term cybernetics, which was first utilized by N. Wiener in 1948 to designate a broad subject area he defined as "control and communication in the animal and the machine." A distinguishing feature of this broad field is the use of feedback information to adapt or steer the entity toward a goal. When this feedback signal is such as to cause changes in the structure or parameters of the system itself, it appears to be self-organizing. *See* ADAPTIVE CONTROL.

Wiener's thinking was influenced by his work during World War II on the aiming of an antiaircraft gun ahead of its target in order to achieve a successful intercept. He was also familiar with the work of physiologists and believed that similar mechanisms are exhibited in biological systems to provide for goal-directed behavior. He developed the statistical methods of autocorrelation, prediction, and filtering of time-series data to provide a mathematical description of both biological and physical phenomena. The use of filtering to remove unwanted information or noise from the feedback signal mimics the selectivity shown in biological systems in which imprecise information from a diversity of sensors can be accommodated so that the goal can still be reached. *See* ESTIMATION THEORY; HOMEOSTASIS; STOCHASTIC CONTROL THEORY.

This similarity between physical and biological systems was also recognized by a number of scientists and engineers in the 1940s, 1950s, and 1960s, such as W. McCulloch, who described the brain as a digital computer. Computing systems that exhibit behavior somewhat akin to that of humans are said to possess artificial intelligence. The field of neural networks is a subset of artificial intelligence in that intelligent behavior can be realized with a collection of computing elements (like biological neurons) connected together to merge various inputs together into a single signal, generally through use of a weighted sum of the individual signals. Usually, a threshold is set and, if this weighted sum is above the set threshold, a neuron fires, indicating that something of importance has happened. Neural networks that adjust the weights by using internal feedback appear to be self-organizing, and are thus consistent with Wiener's view of cybernetic systems. *See* ARTIFICIAL INTELLIGENCE; CONTROL SYSTEMS; ELECTRICAL COMMUNICATIONS; HUMAN-FACTORS ENGINEERING; HUMAN-MACHINE SYSTEMS; NEURAL NETWORK; SYSTEMS ENGINEERING.

Donald W. Bouldin

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Cycadales

An order of the class Cycadopsida of the division Pinophyta (gymnosperms) consisting of four families, Cycadaceae, Stangeriaceae, Zamiaceae, and Boweniaceae, with perhaps 100 species. The order dates from the upper Carboniferous and has few living representatives. The cycads were distributed worldwide in the Mesozoic, but today are restricted to subtropical and tropical regions, with the plants occurring in small colonies; few have broad distributions.

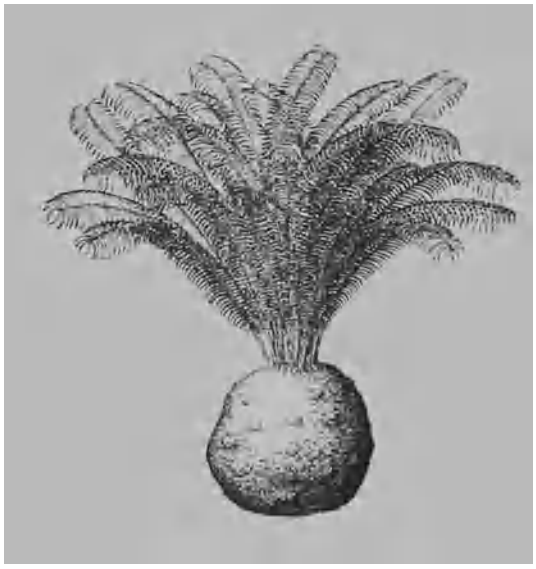
The cycads, often incorrectly referred to as palms, range from a few inches (*Zamia*) to 65 ft (20 m; *Macrozamia*) tall. The stems are cylindrical and often branched; in some the stem is subterranean, in others it is mainly above the ground. The pinnate (or bipinnate in *Bowenia*) leaves are borne at the apex of the stems. Microsporophylls and megasporophylls are borne in cones of highly varied appearance. Male and female cones appear on separate plants (dioecious). *See* CYCADOPSIDA; PINOPHYTA; PLANT KINGDOM.

Thomas A. Zanoni

Cycadeoidales

An order of extinct plants that formed a conspicuous part of the landscape during the Triassic, Jurassic, and Cretaceous periods. The Cycadeoidales (or Bennettitales) had unbranched or sparsely branched trunks with a terminal crown of leaves (family Cycadeoidaceae), or they were branched, at times profusely (family Williamsoniaceae).

The squat, barrel-shaped trunks of members of the Cycadeoidaceae were covered with a dense armor of persistent leaf bases, and were terminated by a crown of pinnately divided leaves (see **illus.**).



Reconstruction of a plant of the genus *Cycadeoidea*, showing terminal crown of leaves, persistent leaf bases, and positions of the fruiting structures. (*Courtesy of T. Delevoryas*)

On the trunk surface, among the leaf bases, were borne complex fruiting structures, each with a central seed-bearing receptacle, surrounded by a whorl of pollen-bearing organs that were derived from modified leaves. Dicotyledonous embryos have been observed in the seeds.

In members of the Williamsoniaceae, fruiting structures were often stalked, and were flowerlike in appearance. They had both pollen-bearing and seed-bearing structures in the same "flower," or these organs may have been borne in separate "flowers." Seeds were borne on the surface of a domelike receptacle, while the pollen-bearing organs were fused at the base to form a whorl of simple or pinnate structures. Foliage was either pinnately divided or entire. No known relatives of the Cycadeoidales exist at the present time, although members of the order Cycadales show some resemblances to the extinct group. *See* CYCADALES; EMBRYOBIONTA; PALEOBOTANY; PLANT KINGDOM. Theodore Delevoryas

Cycadopsida

A class of the division Cycadophyta consisting of a single order, Cycadales, dating back to the Triassic and most abundant in the Jurassic. There are 100–150 living species of cycads in two families, Cycadaceae with one genus and Zamiaceae with ten genera. Most cycads are tropical, but some extend into warm temperate regions. Wild cycads are threatened by collection for horticulture so are protected by international treaties. They were used as traditional foods in some places but contain potent carcinogens and neurotoxins. *See* CYCADALES; GNETALES; WELWITSCHIALES.

Pinnately compound leaves on massive, usually unbranched trunks emerge in flushes, sometimes with the rachis or leaflets coiled like fiddleheads. Large pollen and seed cones crown different individuals. Pollen cones have numerous wedge-shaped scales, each covered below with clustered microsporangia. Seed cones of Zamiaceae are similar, with two seeds on each scale, but megasporophylls of Cycadaceae are leaflike, with two to eight seeds. As in *Ginkgo*, fertilization is by swimming sperm. *See* PINOPHYTA; PLANT KINGDOM. James E. Eckenwalder

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Cyclanthales

An order of flowering plants, division Magnoliophyta (Angiospermae), sometimes called Synanthae or Synanthales, in the subclass Arecidae of the class Liliopsida (monocotyledons). The order consists of the single tropical American family Cyclanthaceae, with about 180 species. They are herbs or, seldom, more or less woody plants, with characteristic leaves that have a basal sheath, a petiole, and an expanded, usually bifid (cleft into two equal parts) blade which is often folded lengthwise (plicate) like that of a palm leaf. The numerous, small, unisexual flowers are crowded into a spadix which serves as a sort of pseudanthium. One of the species, *Carludovicia palmata*, is a principal source of fiber for panama hats. *See* ARECIDAE; LILIOPSIDA; MAGNOLIOPHYTA; PLANT KINGDOM. Arthur Cronquist

Cyclic nucleotides

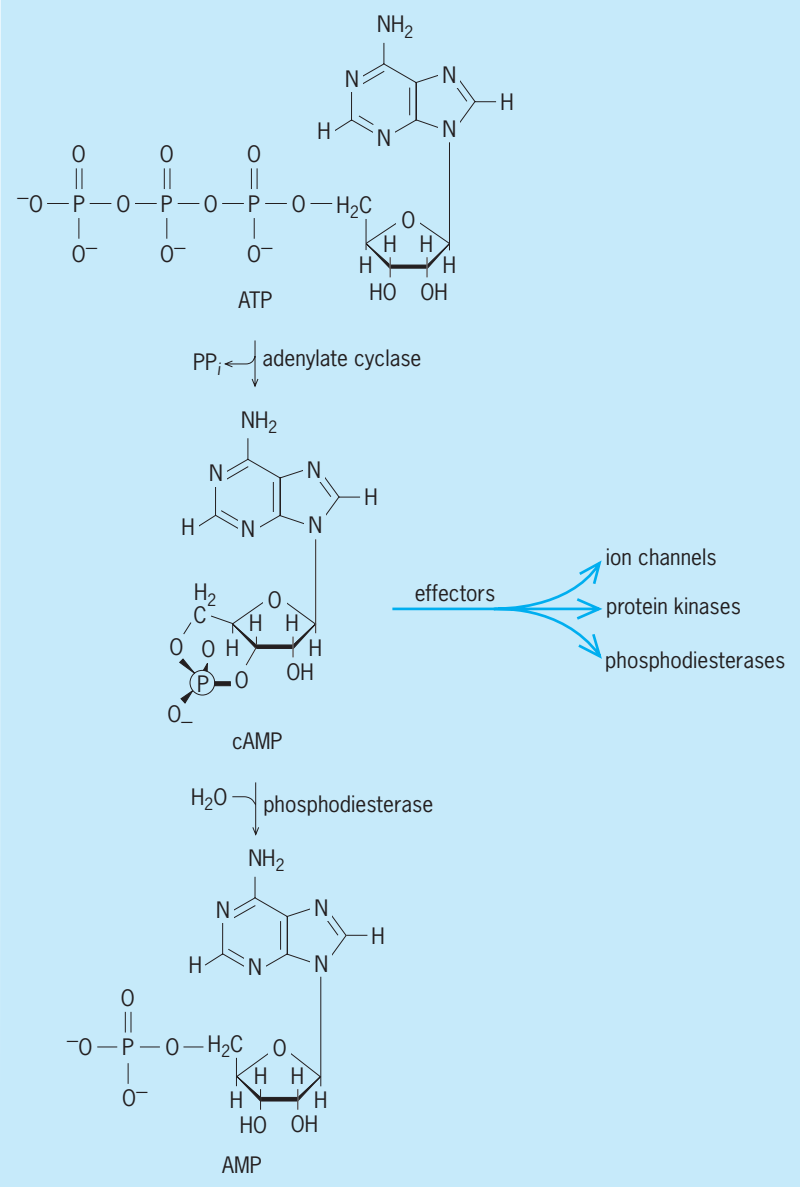
Cyclic nucleotides, derivatives of nucleic acids, control the activity of several proteins within cells to regulate and coordinate metabolism. They are members of a group of molecules known as intracellular second messengers; their levels are regulated by hormones and neurotransmitters, which are the extracellular first messengers in a regulatory pathway. Cyclic nucleotides are found naturally in living cells, from bacteria, to yeast, to human cells.

Formation and degradation. Two major forms of cyclic nucleotides are characterized: 3′,5′-cyclic adenosine monophosphate (cyclic AMP or cAMP) and 3′,5′-cyclic guanosine monophosphate (cyclic GMP or cGMP). Like all nucleotides, cAMP and cGMP contain three functional groups: a nitrogenous aromatic base (adenine or guanine), a sugar (ribose), and a phosphate. Cyclic nucleotides differ from other nucleotides in that the phosphate group is linked to two different hydroxyl (3′ and 5′) groups of the ribose sugar and hence forms a cyclic ring. This cyclic conformation allows cAMP and cGMP to bind to proteins to which other nucleotides cannot.

Specific enzymes are responsible for the formation of cyclic nucleotides. Adenylate cyclase forms cAMP from the nucleotide adenosine triphosphate (ATP), and guanylate cyclase forms cGMP from the nucleotide guanosine triphosphate (GTP) [see **illus.**]. These enzymes form cyclic nucleotides at a very slow rate when cells are in a basal, unstimulated state. However, when certain hormones and neurotransmitters bind to specific receptors on the surface of cells, the cyclases are stimulated more than a hundredfold in their rate of cyclic nucleotide synthesis. The result is a large increase in the intracellular concentration of either cAMP or cGMP.

Cyclic nucleotides are degraded by cyclic nucleotide phosphodiesterases, which cleave the phosphoester bond between the phosphate and the 3′ hydroxyl group of the ribose to generate the 5′-monophosphate nucleotide (see illus.). Many different types of phosphodiesterases are found in various tissues of the body, and several clinically useful drugs inhibit the activity of phosphodiesterases so that cyclic nucleotide concentrations are increased in specific tissues. An example is theophylline, used to treat bronchial asthma.

The increase in cAMP or cGMP triggered by hormones and neurotransmitters can have many different effects on any individual cell. The type of effect is dependent to some extent on the cellular proteins to which the cyclic nucleotides may bind. Three



Biochemical structures of ATP, cAMP, and AMP. In cAMP, a cyclic ring involves the 3′ and 5′ hydroxyl groups of the ribose with the phosphate. There are three major effector mechanisms for the second messenger action of cAMP in human cells.

types of effector proteins are able to bind cyclic nucleotides.

Effector proteins. Protein kinases constitute the first type of effector protein regulated by cyclic nucleotides. They are enzymes which are able to transfer a phosphate group to (phosphorylate) individual amino acids of other proteins. This action often changes the function of the phosphorylated protein. For example, the enzyme phosphorylase is activated by phosphorylation, resulting in the release of glucose into the bloodstream from glycogen stored in the liver. Cyclic nucleotides activate the cAMP-dependent protein kinases (PKA) and the cGMP-dependent protein kinases (PKG). PKA is found in all cell types, where it regulates the flow of ions across the cell membrane, the activity of metabolic enzymes such as phosphorylase, the growth and division of cells, and the rate at which individual proteins are synthesized. PKA phosphorylates distinct cellular proteins involved in each of these aspects of cell function. PKG is found only in certain cells of the body, including smooth muscle cells lining the blood vessels where it regulates blood flow, and neurons of the brain where it regulates their response to neurotransmitters.

The second type of protein regulated by cyclic nucleotides comprises ion channels. These proteins are found in the outer plasma membrane of some cells, and the binding of cyclic nucleotides can alter the flow of sodium ions across the cell membranes. These ion channels are found in greatest abundance in the sensory neurons of the retina and nose as well in some regions of the brain. The binding of cyclic nucleotides to these ion channels alters the sensitivity of the neurons to neurotransmitters. In the retina, cGMP plays a critical role in vision by regulating an ion channel.

The third type of protein to which cyclic nucleotides bind are the cyclic nucleotide phosphodiesterases. These enzymes are responsible for the degradation of cyclic nucleotides. The degradation rate can be affected by the amount and nature of the cyclic nucleotides present in a cell. For example, high levels of cGMP can inhibit the phosphodiesterase that is responsible for the degradation of cAMP.

In bacteria, cAMP can bind to a fourth type of protein, which can also bind to deoxyribonucleic acid (DNA). This catabolite gene activator protein (CAP) binds to specific bacterial DNA sequences, stimulating the rate at which DNA is copied into ribonucleic acid (RNA) and increasing the amount of key metabolic enzymes in the bacteria. This type of regulation does not occur in eukaryotic or animal cells.

In humans, cyclic nucleotides acting as second messengers play a key role in many vital processes and some diseases. In the brain, cAMP and possibly cGMP are critical in the formation of both long-term and short-term memory. In the liver, cAMP coordinates the function of many metabolic enzymes to control the level of glucose and other nutrients in the bloodstream. Cyclic GMP mediates the effects of nitroglycerin on smooth muscle cells in the arteries of the heart to alleviate the lack of blood flow associated with angina. Cholera results in abnormally high levels of cAMP in the intestinal epithelial cells and the loss of large amounts of fluid, resulting in diarrhea. In some cases of cystic fibrosis, the inability of cAMP to regulate an ion channel causes the disease. *See* ADENOSINE TRIPHOSPHATE (ATP); ENZYME; NUCLEIC ACID; NUCLEOPROTEIN; NUCLEOTIDE; PROTEIN.

Michael D. Uhler

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Cycliophora

A recently described phylum in the animal kingdom. Only one microscopic species, *Symbion pandora*, has been described. Its sessile stage is approximately 0.3 mm in length and 0.1mm wide. Surprisingly, Cycliophora was not discovered in an exotic biotope such as a deep-sea vent or marine caves, but on the Norwegian lobster. All known occurrences are from the crustacean mouth limbs, which can be totally encrusted with thousands of the sessile stage of the microscopic animal (**Fig. 1**). Two other undescribed species, very similar to the type species, are found on the American lobster (*Homarus americanus*) and the common lobster (*H. gammarus*).

The sessile animal has a circular mouth surrounded by a ring of cilia (**Fig. 2**). The mouth ring is used for filtering small food particles, such as bacteria or algae. The anatomy is relatively simple, with a U-shaped gut, similar to what is found in bryozoans or some sessile rotiferans. It is the extremely complex life cycle of *S. pandora* that makes the phylum unique. The life cycle consists of an asexual and a sexual generation with no less than two types of free-swimming larva, dwarf male and female; two stages of sessile feeding animals, which brood the male and female; and one type of larva inside a brooding chamber (called a marsupium). Furthermore, the sessile animal has internal budding, whereby it grows by loss of the head itself and replacement of the old gut and feeding system with a new bud coming from embryonic cells in the basal part of the body. The larva developed inside the feeding animal is called the Pandora larva. The feeding animal thus has developed both a new bud with a head and the Pandora larva. However, inside the posterior segment of the Pandora larva, a new internal bud with the head of a new feeding animal is developed before the larva leaves the sessile mother animal.

When the Pandora larva emerges, it settles on the setae of the lobster very close to the maternal individual, and the internal bud pops up from the rear to form a new feeding-stage animal. In this way the *S. pandora* can multiply very fast and completely cover the mouth limbs in a whitish cast. After several cycles, or before the lobster molts or dies, the feeding stage produces a single male or female in the marsupium in the same way as the Pandora larva. It is thought that the molting hormones of the lobster may trigger the sexual life cycle.

The dwarf male is released from the marsupium with an undeveloped testis and without a penis. It settles on a feeding animal. If the feeding animal is brooding a female inside, the newly settled male begins to make internal buddings and it forms up to three testes and penial structures. The female looks very similar to a Pandora larva except that she lacks the internal budding; rather she has a single large egg posteriorly. This egg is fertilized inside the female before release from the marsupium. The fertilized egg (zygote) immediately begins to cleave inside the female when it escapes the brooding chamber. She settles very close to her mother on the same host and now broods a totally new type of larva, the chordoid larva inside herself. In fact, the larval tissue engulfs all her tissue, so that only her cuticle remains. The chordoid larva hatches from the female

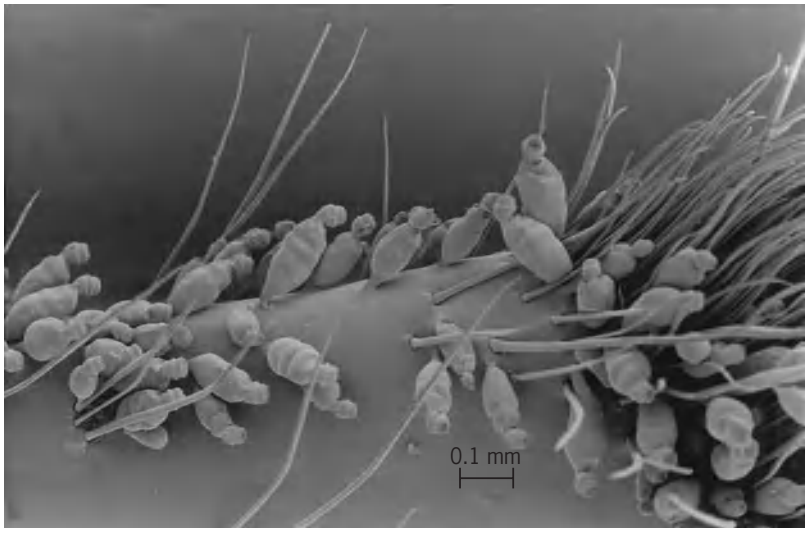

Fig. 1. Scanning electron micrograph of the Norwegian lobster's mouthpart with numerous feeding stages of *Symbion pandora*. (*Courtesy of P. Funch and R. M. Kristensen*)

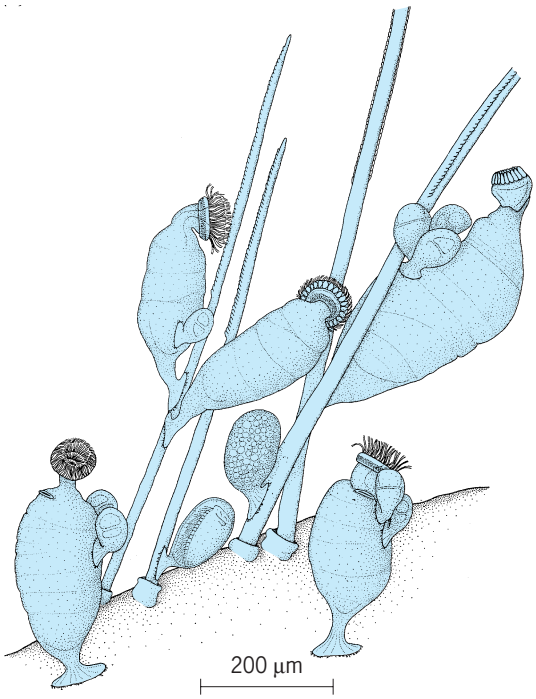

Fig. 2. Drawing of the first described cycliophoran, *Symbion pandora*, from the Norwegian lobster, *Nephrops norvegicus*. (*Drawn by Stine Elle*)

exuvium head first and swims away. The chordoid larva is the dispersal stage. It is a good swimmer and has a strong chorda-like structure similar to that found in the larva of tunicates, although it is not believed that these structures are homologous in the two groups. The chordoid larva settles on a new host and forms a new feeding animal; a new life cycle can begin.

Most of the life cycle of *S. pandora* was elucidated in the Faeroe Islands in the North Atlantic, but there is no reason why the Cycliophora should not be found on other invertebrates worldwide.

Reinhardt Møbjerg Kristensen

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Cycloamylose

Any of a group of cyclic oligomers of glucose in the normal C-1 conformation in which the individual glucose units are connected by 1,4 bonds. They are called Schardinger dextrins (after the discoverer), cyclodextrins (to emphasize their cyclic character), or cycloamyloses (to emphasize both their cyclic character and their amylose origin). The most common ones are cyclohexaamylose (with six glucose units in a cyclic array) and cycloheptaamylose (with seven glucose units in a cyclic array). The toroidal shape of these molecules has been determined by x-ray analysis (**Fig. 1**). The molecular weights are around 1000.

In water or in a mixture of water and dimethyl sulfoxide, cycloamyloses form 1:1 complexes with many organic molecules and most benzenoid derivatives. X-ray and spectroscopic evidence indicates that the guest molecule (usually a small organic molecule) is bound tightly in the cavity of the host molecule (the cycloamylose) just like a key fitting into a lock. The result is an inclusion complex, which is what is formed by enzymes when they first bind the molecules whose reactions they subsequently catalyze. *See* CATALYSIS; CLATHRATE COMPOUNDS.

Inclusion complex formation is probably the most important similarity between cycloamyloses and enzymes. In the cycloheptaamylose-catalyzed hydrolysis of *p*-nitrophenyl acetate, a nonlinear increase in the rate of reaction is obtained when one increases the concentration of cycloamylose. However, a linear increase is obtained when the same data are plotted in a double reciprocal manner. These two concentration dependencies are exactly those seen in enzyme-catalyzed reactions, and show that enzymes and cycloamyloses increase the rate of reaction in the same way.

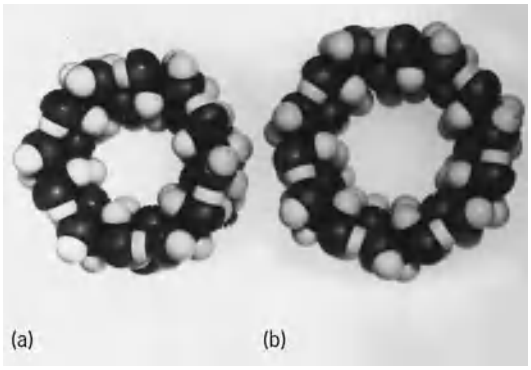

Fig. 1. Molecular models of cycloamyloses viewed from the secondary hydroxyl side of the torus. (*a*) Cyclohexaamylose. (*b*) Cycloheptaamylose. (*From M. L. Bender and M. Komiyama, Cyclodextrin Chemistry, Springer, 1978*)

Specificity is one of the most important aspects of enzyme action. A particular enzyme can catalyze a given reaction but cannot catalyze a closely related reaction. The cycloamyloses provide the answer to how an enzyme recognizes a given compound but ignores a closely related compound.

The importance of the stereochemistry of binding may be seen easily in the models of complexes of corresponding meta- and para-substituted esters with cyclohexaamylose: cyclohexaamylose *m-t*-butylphenyl acetate and cyclohexaamylose *p-t*-butylphenyl acetate. In the former complex, a glucosidic hydroxyl group is practically adjacent to the ester carbonyl group with which it is to react, leading to a facile reaction (**Fig. 2***a*). In the latter complex, the complex, the glucosidic hydroxyl group is several tenths of a nanometer away from the ester carbonyl group

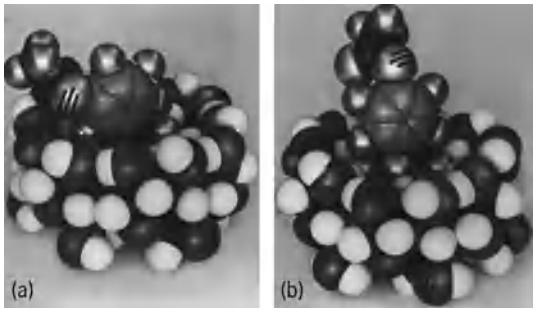

Fig. 2. Molecular models of cycloamylose complexes. (*a*) Cyclohexaamylose-*m-t*-butylphenyl acetate complex. (*b*) Cyclohexaamylose-*p-t*-butylphenyl acetate complex. (*From R. L. Van Etten et al., Acceleration of phenyl ester cleavage by cycloamyloses: A model for enzymatic specificity, J. Amer. Chem. Soc., 89:3249, 1967*)

with which it must react, leading to a slow reaction (Fig. 2*b*).

Cycloamyloses have been used to mimic many enzymes, such as ribonuclease, transaminase, and carbonic anhydrase. The catalysis of the hydrolysis of phenyl acetates discussed above is a mimicry of the enzyme chymotrypsin by a cycloamylose. Both cycloamyloses and chymotrypsin catalyze via three steps: binding; acylation of the catalyst; and deacylation of the acyl-catalyst. The mechanisms by which cycloamyloses and chymotrypsin catalyze this reaction are identical: cycloamyloses form a noncovalent complex followed by the formation of an acyl-cycloamylose intermediate, while chymotrypsin forms a noncovalent enzyme-substrate complex followed by an acyl-chymotrypsin intermediate. Cycloamyloses, like chymotrypsin, catalyze the hydrolysis of both esters and amides.

Cycloamyloses have also been used to effect selectivity in several chemical reactions. Anisole can be chlorinated by HOCl in the presence of the cyclohexaamylose exclusively in the position para to the methoxy group, although the absence of cyclohexaamylose gives the usual mixture of ortho and para chlorinated products. The ortho position of anisole is buried inside the cavity when the anisole is bound in the cyclohexaamylose, and so reaction takes place in the para position only, although the methoxy group is an ortho, para director in aromatic electrophilic substitution. In a similar sense, it was found that the introduction of aldehyde groups into many phenols by chloroform and sodium hydroxide leads solely to the para-substituted benzaldehyde because the dichlorocarbene attacking agent is prevented from approaching the ortho position by the cycloheptaamylose.

Cycloamyloses are the premier example of chemical compounds that act like enzymes in inclusion complex formation followed by reaction. *See* ENZYME; STEREOCHEMISTRY; STEREOSPECIFIC CATALYST; STERIC EFFECT (CHEMISTRY).

Myron L. Bender

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Cyclocystoidea

An extinct class of small, disk-shaped echinozoans in which the lower surface of the body probably consisted of a suction cap for adhering to substrate, and the upper surface was covered by separate plates arranged in concentric rings. The mouth lay at the center of the upper surface, and the anus, also on the upper surface, lay some distance between the margin and the mouth. Little is known of the habits of cyclocystoids. They occur in rocks of middle Ordovician to middle Devonian age in Europe and North America. *See* ECHINODERMATA; ECHINOZOA.

Howard B. Fell

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Cycloid

A curve traced in the plane by a point on a circle that rolls, without slipping, on a line. If the line is the x axis of a rectangular coordinate system, at whose origin O the moving point P touches the axis, parametric equations of the cycloid are $x = a(\theta - \sin\theta)$, $y = a(1 - \cos\theta)$, when a is the radius of the rolling circle, and the parameter θ is the variable angle through which the circle rolls (see **illus.**). One arch is obtained where θ assumes all values from 0 to 2π. The length of an arch is $8a$, and the area bounded by an arch and the x axis is $3\pi a^2$ (three times the area of the rolling circle).

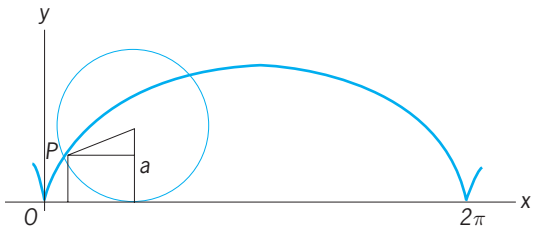

Diagram of cycloid; symbols explained in text.

Two important properties of the cycloid are the following: (1) It is isochronous, that is, if the line is horizontal, then the time of descent of a particle starting from rest at a point of a cycloidal runway and falling to its lowest point is independent of the starting point. (2) It is a brachistochrone; that is, if P_1, P_2 are two points in a vertical plane, but not in a vertical line, the curve down which a particle will fall from P_1 to P_2 in the shortest time is an arc of a cycloid. *See* ANALYTIC GEOMETRY.

Leonard M. Blumenthal

Cyclone

A vortex in which the sense of rotation of the wind about the local vertical is the same as that of the Earth's rotation. Thus, a cyclone rotates clockwise in the Southern Hemisphere and counterclockwise in the Northern Hemisphere. In meteorology the term cyclone is reserved for circulation systems with horizontal dimensions of hundreds of kilometers (tropical cyclones) or thousands of kilometers (extratropical cyclones). For such systems the Coriolis force due to the Earth's rotation and the centrifugal force due to flow curvature are both directed to the right of the direction of the flow. Away from the surface of the Earth these two forces are approximately balanced by the pressure gradient force, which is directed toward low pressure (**Fig. 1**), so that there

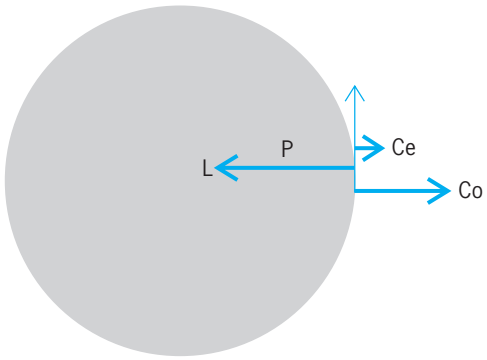

Fig. 1. Force balance for gradient flow about a cyclone in the Northern Hemisphere. Thin arrow indicates wind velocity; L = low-pressure center; Co = Coriolis force; Ce = centrifugal force; and P = pressure gradient force.

must be a pressure minimum at the center of the cyclone, and cyclones are sometimes simply called lows.

This three-way balance is referred to as the gradient wind. In large-scale extratropical cyclones the centrifugal force is of secondary importance. The balance is then primarily between the Coriolis force and the pressure gradient force. In this case the flow, which is parallel to the isobars with high pressure to the right of the direction of motion (in the Northern Hemisphere), is said to be in geostrophic balance. *See* AIR PRESSURE; CORIOLIS ACCELERATION; GRADIENT WIND; VORTEX.

Extratropical cyclones. These are the common weather disturbances which travel around the world from west to east in mid-latitudes. They are generally associated with fronts, which are zones of rapid transition in temperature. In a typical midlatitude cyclone (**Fig. 2**) the air poleward of the

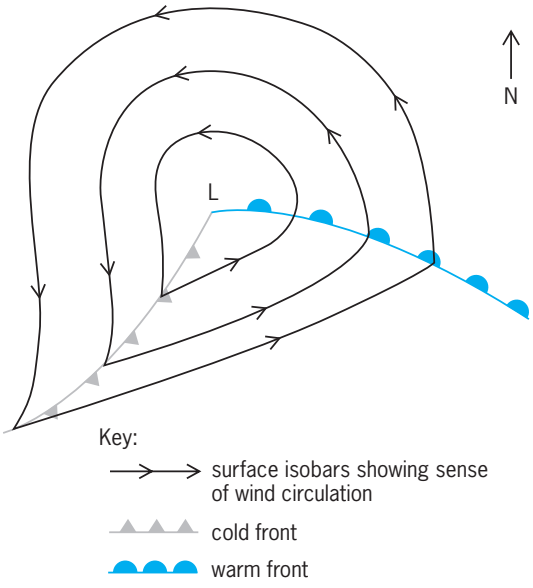

Fig. 2. Plane view of an extratropical cyclone in the Northern Hemisphere as depicted on a surface weather map. The L designates a pressure low.

frontal boundary is cool and dry, while the air equatorward is warm and moist. The poleward and upward motion of warm moist air south of the warm front, which extends eastward from the low-pressure center, is responsible for much of the precipitation associated with cyclonic storms. The equatorward and downward motion of cool dry air behind the cold front, which extends southwest of the low-pressure center, is responsible for the fine weather which generally follows the passage of a midlatitude cyclone.

Formation of extratropical cyclones is referred to as cyclogenesis. In the classic Norwegian cyclone model, which is favored by many weather forecasters, cyclogenesis begins with the development of a wavelike disturbance on a preexisting frontal discontinuity separating a cold polar air mass from a warm subtropical air mass (**Fig. 3***a*). The circulation associated with the wave then distorts the frontal surface to produce a closed vortical circulation with a warm front ahead of the vortex and a trailing cold front (Fig. 3*b*). In the later stages the cold front may overtake the warm front to produce an occlusion, a region where at the surface the warm air is limited to a narrow band (Fig. 3*c*).

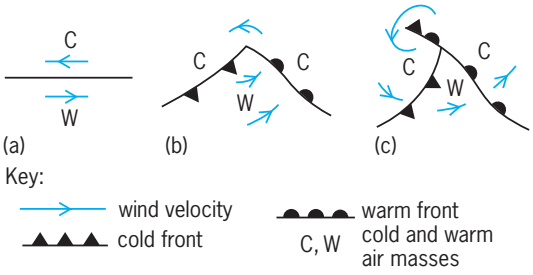

Fig. 3. Classic Norwegian cyclone model showing three stages in the life of a polar front cyclone. (*a*) Frontal discontinuity, (*b*) distorted frontal surface, (*c*) occlusion. (*After B. J. Hoskins, Dynamical processes in the atmosphere and the use of models, Quart. J. Roy. Meteorol. Soc., 109:1–21, 1983*)

A more modern theoretical model of cyclogenesis, on the other hand, suggests that most extratropical cyclones develop as dynamical instabilities in broad regions of enhanced horizontal temperature gradients and vertical shear of the geostrophic wind associated with the subtropical jet stream. In this instability process, which is referred to as baroclinic instability, the cyclone grows by converting potential energy associated with the horizontal temperature gradient in the mean flow into kinetic energy. As baroclinically unstable systems develop, the disturbance motion tends to intensify temperature contrasts near the surface so that warm and cold fronts develop. Thus, in this theory, fronts are considered to be secondary features produced by cyclogenesis, rather than acting as causes.

Cyclogenesis occurs over very broad areas of the Earth. However, it is most frequent over the western oceans (such as in the regions east of Japan and off

the east coast of North America) and in the lee of major mountain barriers (such as the Rockies and the Alps).

Occasionally, over maritime regions during the cold season, extratropical cyclones can develop rapidly into dangerously intense storms with winds comparable to those found in hurricanes. If decreases in surface pressure occur at a rate exceeding 1 millibar (100 pascals) per hour for a period of greater than 24 h, the development is called explosive cyclogenesis. Explosive cyclogenesis tends to occur in strong baroclinic zones when a weak surface disturbance receives reinforcement from a cyclonic disturbance aloft, and the static stability of the atmosphere (that is, the resistance to vertical motions) is weak so that strong cumulus convection occurs. *See* BAROCLINIC FIELD; FRONT; GEOSTROPHIC WIND; JET STREAM.

Tropical cyclones. By contrast, tropical cyclones derive their energy from the release of latent heat of condensation in precipitating cumulus clouds. The moisture necessary to sustain this precipitation is supplied by the frictionally driven flow of moist air toward low pressure near the Earth's surface and its subsequent ascent into the cumulus clouds. Over the tropical oceans, where moisture is plentiful, tropical cyclones can develop into intense vortical storms (hurricanes and typhoons), which can have wind speeds in excess of 200 mi/h (100 $m \cdot s^{-1}$). *See* HURRICANE; PRECIPITATION (METEOROLOGY); STORM; TROPICAL METEOROLOGY; WIND.

J. R. Holton

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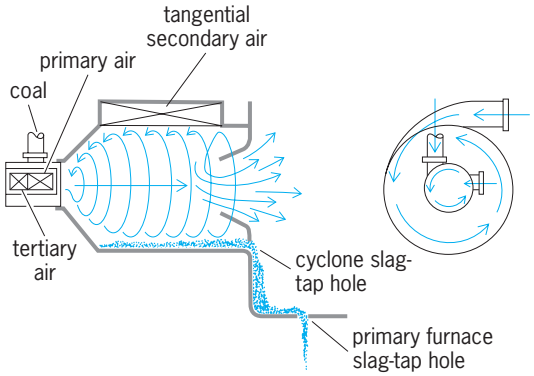
Cyclone furnace

A water-cooled horizontal cylinder in which fuel is fired and heat is released at extremely high rates. When firing coal, the crushed coal, approximately 95% of which is sized at 14 in. (0.6 cm) or less, is introduced tangentially into the burner at the front end of the cyclone (see **illus.**). About 15% of the combustion air is used as primary and tertiary air to impart a whirling motion to the particles of coal. The whirling, or centrifugal, action on the fuel is further increased by the tangential admission of high-velocity secondary air into the cyclone. *See* FURNACE CONSTRUCTION.

The combustible is burned from the fuel in the cyclone furnace, and the resulting high-temperature gases melt the ash into liquid slag, a thin layer of which adheres to the walls of the cyclone. The incoming fuel particles, except those fines burned in suspension, are thrown to the walls by centrifugal force and caught in the running slag. The secondary air sweeps past the coal particles embedded in the slag surface at high speed. Thus, the air required to burn the coal is quickly supplied and the products of combustion are rapidly removed.



Schematic diagram of cyclone furnace. (*After E. A. Avallone and T. Baumeister III, eds., Marks' Standard Handbook for Mechanical Engineers, 9th ed., McGraw-Hill, 1987*)

The products of combustion are discharged through a water-cooled reentrant throat at the rear of the cyclone into the boiler furnace. The part of the molten slag that does not adhere to the cyclone walls also flows toward the rear and is discharged through a taphole into the boiler furnace.

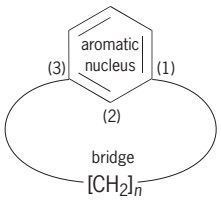
Essentially, the fundamental difference between cyclone furnaces and pulverized coal–fired furnaces is the manner in which combustion takes place. In pulverized coal–fired furnaces, particles of coal move along with the gas stream; consequently, relatively large furnaces are required to complete the combustion of the suspended fuel. With cyclonic firing, the coal is held in the cyclone and the air is passed over the fuel. Thus, large quantities of fuel can be fired and combustion completed in a relatively small volume, and the boiler furnace is used to cool the products of combustion.

Gas and oil can also be burned in cyclone furnaces at ratings and with performances equaling those of coal firing. When oil is the fuel, it is injected adjacent to the secondary air ports and directed downward into the secondary airstream. The oil is picked up and sufficiently atomized by the high-velocity air. Gas is fired similarly through flat, open-ended ports located in the secondary air entrance. *See* BOILER; STEAM-GENERATING FURNACE; STEAM-GENERATING UNIT.

G. W. Kessler

Cyclophane

A molecule composed of two building blocks: an aromatic ring (most frequently a benzene ring) and an aliphatic unit forming a bridge between two (or more) positions of the aromatic ring. Structure (**1**),

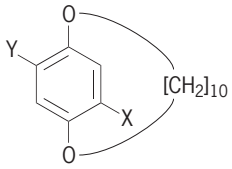


(**1**)

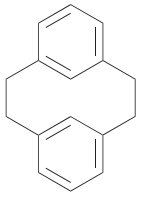
an [*n*](1,3)cyclophane, provides a general example for a simple cyclophane and its nomenclature. This cyclophane is also known as [*n*]metacyclophane. In structure (**1**), *n* is the number of $[CH_2]$-groups in the bridge, and (1) and (3) are the starting positions of the bridge.

Nomenclature. Those cyclophanes that contain heteroatoms in the aromatic part are generally called phanes, as "cyclo" in cyclophanes more strictly means that only benzene rings are present in addition to aliphatic bridges.

Phane molecules containing hetero atoms in the aromatic ring (for example, nitrogen or sulfur) are called heterophanes, while in heteraphanes they are part of the aliphatic bridge [structures (**2**) and (**3**)]. By coupling meta-xylylene dibromide with sodium (Wurtz coupling) the [2.2]metacyclophane (**4**) is ob-



(**2**) X = Y = H
(**3**) X = CO_2H, Y = Br

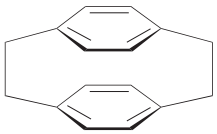


(**4**)

tained. The name is formed by putting the lengths (2 atoms) of the 2 bridges in square brackets, and the (1) and (3) positions at the benzene ring where the bridge starts and ends, in brackets. As the 1,3 positions are also called meta, this hydrocarbon (**4**) is usually named [2.2] metacyclophane. *See* AROMATIC HYDROCARBON.

The first heteracyclic [*n*]paracyclophane (**2**) that was synthesized, the [12]paracyclophane, is called ansa compound [ansa (Latin), handle] because of its geometry. The tight embracing of structure (**2**) has stereochemical significance. Compound (**3**), with its hindered free rotation of the aliphatic bridge, exists as atropisomers (axial chirality) and can be separated into its stable enantiomers.

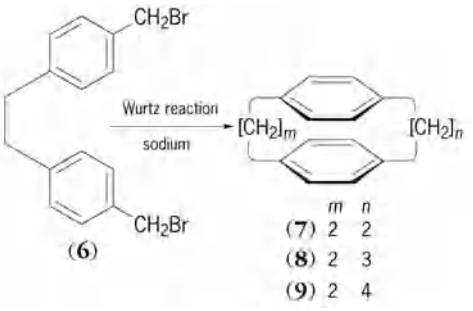
Synthesis. Modern syntheses of cyclophanes began with the first synthesis of [2.2]paracyclophane (**5**). The hydrocarbon was obtained by extraction



(**5**)

from pyrolysis products of *p*-xylene.

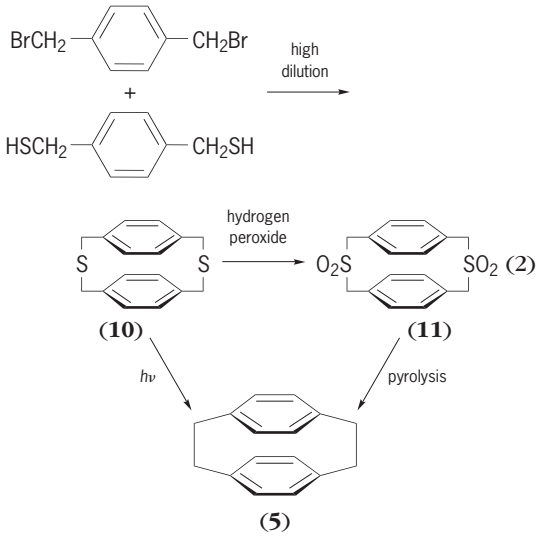
The second major advance in cyclophane chemistry resulted from the work of D. J. Cram, the first directed synthesis of [2.2], [3.2], and [4.2]paracyclophanes (**7**)–(**9**) that used a high-dilution intramolecular Wurtz reaction with molten sodium in xylene starting from bis(brommethyl)bibenzyl (**6**), as shown in reaction (1). The Wurtz coupling is a radical re-



(1)

action induced by molten sodium forming carbon-carbon single bonds starting from alkyl halides (for example, **6**). It is necessary to use a low concentration of starting materials in a high volume of solvent (high dilution)to suppress intermolecular side reactions and to obtain mainly the intramolecular coupling shown in reaction (1).

In the 1970s, new synthetic methods of forming highly strained cyclophanes became available. One type, sulphone pyrolysis (also known as sulfur dioxide extrusion), can be applied to the synthesis of many cyclophanes, and it has become one of the most widely applicable routes [reaction (2)]. By



(**5**)

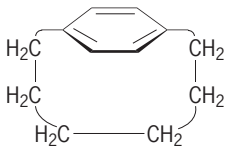
eliminating sulfur photochemically or by pyrolysis, strain is introduced into the ring system starting from molecules like (**10**) and (**11**). Compound (**5**) can be obtained in high yield by this synthetic approach. *See* ORGANIC SYNTHESIS.

Properties. Cyclophanes, with their remarkable structure, often exhibit peculiarities in electronic and spectroscopic properties, the investigation

of which is of fundamental scientific value. For example: Cyclophanes usually are not at all planar molecules; they exhibit an interesting stereochemistry (arrangement of atoms in three-dimensional space); molecular parts are placed and sometimes fixed in unusual orientations toward each other, and often the molecules are not rigid but (conformationally) flexible. Ring strain and, as a consequence, deformation of the benzene rings out of planarity are often encountered. Electronic interactions between aromatic rings fixed face to face [like the two benzene rings in (**5**)] can take place. In addition, influence on substitution reactions in the aromatic rings results; that is, substitutents in one ring [for example, of (**5**)] induce transannular electronic effects in the other ring, often leading to unexpected products. *See* CONFORMATIONAL ANALYSIS; STEREOCHEMISTRY.

Aromatic rings in many cyclophanes are not planar but distorted through bending (referred to as bent and battered benzene rings). It has been shown by x-ray crystal structure analysis that benzene rings can be distorted into boat, chair, and twist forms by clamping them in cyclophane structures by short bridges. These stereochemical aberrations result in spectroscopic and chemical consequences that are of fundamental interest to all chemists, particularly in connection with aromatic units such as benzene, naphthalene, azulene, or pyridine. *See* BENZENE.

Tightly clamped cyclophanes are usually highly strained. The ring strain reaches an extreme in highly deformed molecules such as [6]paracyclophane (**12**). This molecule is the smallest stable



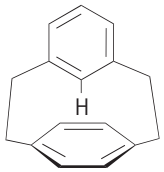
(**12**)

member of [*n*]paracyclophanes that can be isolated. In ring-substituted [6]paracyclophanes, it was found by x-ray crystal structure analysis that the distortion angle is up to 22.4° and the bond angle in the aliphatic bridge was expanded to 126.5° (the ideal value in an unstrained molecule is 109.5°). Some of the bridging protons in the [6]paracyclophane are located above the plane of the aromatic ring and therefore shifted strongly upfield to δ[ppm]: −0.6. In ^{1}H-NMR spectroscopy, the aromatic protons generally appear in a region of δ[ppm]: 7–9. Some bridging protons of the [6]paracyclophane are oriented above the aromatic ring, and as a consequence these protons are strongly shifted upfield to high magnetic field strength to δ[ppm]: −0.6 (beyond the internal standard substance tetramethylsilane, which by definition equals 0) influenced by the aromatic π system. Cyclophane chemistry provides the opportunity to gradually increase the distortions and to measure the resulting effects. *See* X-RAY CRYSTALLOGRAPHY.

Benzene rings can be arranged into multilayer cyclophanes, and transannular electronic effects through more than two aromatic rings can be detected spectroscopically. Furthermore, cyclophane molecules can orient themselves as planar-chiral and helical arrangements; and the resulting enantiomers, once separated, can be examined with respect to their properties toward circularly polarized light.

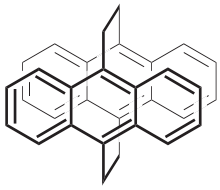
Cyclophane chemistry, with the three-dimensional geometry of its molecules, has revealed many unusual molecules. It is the wide variety of arrangements of partial structures, fixed or conformationally flexible, that affords such a rich field for research into molecular design, synthesis, structural analysis, physical investigations, chemical reactions, and all the way to supramolecular chemistry. *See* DENDRITIC MACROMOLECULE.

Configuration. Cyclophanes have been synthesized in both fairly simple structures and more elaborate structures. An interesting representative of the cyclophane family is the rather simple hydrocarbon [2.2]metaparacyclophane (**13**), which shows a dy-



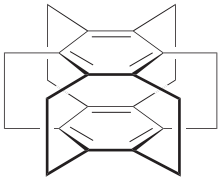
(**13**)

namic flexibility of the meta-substituted benzene ring that can be studied by nuclear magnetic resonance spectroscopy. The compound [2.2]anthracenophane (**14**) can easily undergo a smooth, photo-



(**14**)

chemically induced cyclization to yield a cage compound. Completely bridged aromatic units lead to [2_6](1,2,3,4,5,6)cyclophane (**15**), also called super-

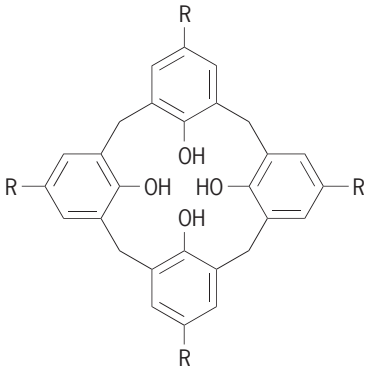


(**15**)

phane, which was synthesized by stepwise formation of the bridges. In sharp contrast to the strained

cyclophanes, which are conspicuous for their face-to-face and bent-and-battered benzene rings, the two benzene rings in superphane (**15**) are planar and undistorted.

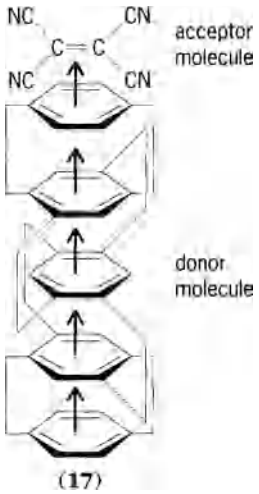
Interest in calixarenes like structure (**16**) stems



(**16**)

from their basketlike shape being adapted to host-guest or receptor-substrate chemistry and the fact that they are easily synthesized. Some members are commercially available, and have been shown to function as hydrolysis catalysts and oxygen carriers. They were named calixarenes because of the chalice-like shape of the molecules and the aromatic (arene) components of the macrocycle. *See* NUCLEAR MAGNETIC RESONANCE (NMR).

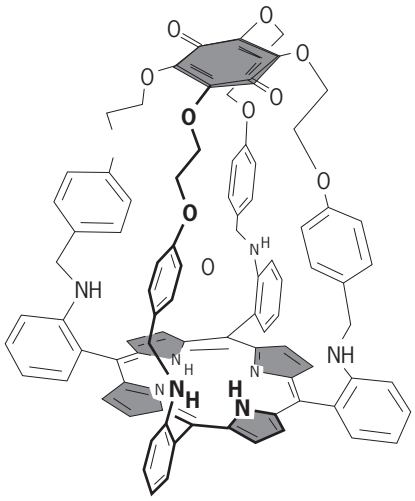
As the number of benzene ring layers is increased is multilayered cyclophanes, the ultraviolet absorption of the donor-acceptor complexes is shifted to longer wavelengths, indicating an increasing donor character of the cyclophane. These spectral characteristics are explained as arising from electronic interaction between the benzene rings. With increasing donor character, the stability of the charge-transfer complex increases. This enhanced π-donor effect is still noticeable in the quintuple-layered cyclophane complex (**17**), with tetracyanoethene as



(**17**)

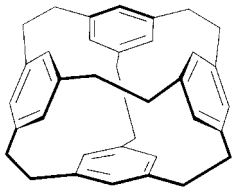
the acceptor molecule. There is no more additional difference seen in the transannular electron-donor capacity in the corresponding sextuple-layered cyclophane. *See* CHEMICAL BONDING.

The quinone-capped porphyrinophane (**18**) is of



(**18**)

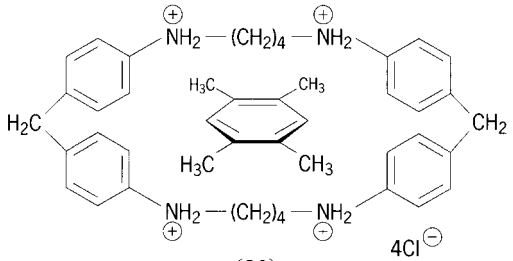
interest as a model compound in connection with biological charge-transfer interactions and provides an example for surprisingly high yields in macrocyclization reactions involving rigid subunits and reversible reaction types. The spheriphane (**19**; $C_{36}H_{36}$), with



(**19**)

its four tetrahedrally arranged benzene rings, constitutes a concave hydrocarbon and a cage molecule with its possibility for encapsulating and transporting small molecules or ions. *See* CAGE HYDROCARBON.

Molecular recognition. Cyclophanes have become important in host-guest chemistry and supramolecular chemistry as they constitute host compounds for guest particles because of their cavities. Recognition processes at the molecular level are understood to mean the ability to design host molecules to encompass or attach selectively to smaller, sterically complementary guests in solution—by analogy to biological receptors and enzymes. For this reason, water-soluble cyclophanes like host (**20**) were synthesized; they have the ability to include nonpolar guest molecules (for example, tetramethylbenzene) in aqueous solution. The stoichiometric inclusion was proven by x-ray crystal structure analysis. *See* MOLECULAR RECOGNITION; POLAR MOLECULE.



(20)

Applications. Cyclophanes also exist in nature as alkaloids, cytotoxic agents, antibiotics, and many cyclopeptides. With the possibility of locating groups precisely in space, cyclophane chemistry has provided building units for molecular niches, nests, hollow cavities, multifloor structures, helices, macropolycycles, macro-hollow tubes, novel ligand systems, and so on. Cyclophane chemistry has become a major component of supramolecular chemistry, molecular recognition, models for intercalation, building blocks for organic catalysts, receptor models, and variation of crown ethers and cryptands. Apart from the various research possibilities and their role as cyclo-oligomers for high-molecular-weight compounds, some cyclophanes find application in the high-tech field. Examples are the [2.2.2]paracyclophane, a cyclophane that exhibits a remarkable high and selective extraction capability for silver ions, which can be extracted from a water phase into a lipophilic solvent; and [2.2]paracyclophane, which has been used for the mechanistic investigation of catalytic hydrogenation, a very important industrial process. *See* MACROCYCLIC COMPOUND; SUPRAMOLECULAR CHEMISTRY.

F. Vögtle; M. Böhme

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Cyclopoida

An order of small copepod Crustacea. The abundant fresh-water and salt-water species form an important link in the food chains of aquatic life, consuming tiny plants, animals, and detritus and, in turn, furnishing food for small fish, some large fish, and other organisms. Some species are important intermediate hosts for human parasites. Females of the family Ergasilidae are parasitic on aquatic animals, while the males are free-swimming. *See* FOOD WEB.

The front part of the body is oval and sharply separated from the tubular hind end, which bears two caudal rami with distinctly unequal setae (**Fig. 1**). Usually 10 body segments are present in the male, and 9 in the female because of fusion of two to form a genital segment. Appendages (**Fig. 2**) are the first and second antennae, mandibles, first and second maxillae, maxillipeds, and five pairs of swimming legs, the last of which are alike and rudimentary. Males usually have rudimentary sixth legs. Appendages are usually biramous, with a 1- or 2-segmented basal portion. One branch is missing from the first antennae, and usually from the second antennae, the mandibles, and fifth legs. The first antennae are never much longer than the anterior portion of the body and consist of 6–17 segments. Usually there are either 12 or 17. Both first antennae of the males are modified for grasping females. The digestive tract consists of a narrow, dilatable esophagus, an expanded stomach, a long intestine, and a short rectum opening on the dorsal side of the last body segment.

A heart is absent. Colorless blood containing ameboid corpuscles is circulated by general body movements and rhythmic movements of the digestive tract. External respiration is through the body surface. Striated muscles are highly developed. The genital systems are shown in Fig. 1. In mating, the male grasps the female with the first antennae and by means of the swimming legs places two kidney-shaped spermatophores at the opening of the female's seminal receptacle. Egg sacs are paired and lateral. Rarely, they are subdorsally located. There are six nauplius and six copepodid stages, the last being the adult. Some species encyst in the copepodid stages to survive adverse conditions. A small brain in the head is connected, around each side of the esophagus, to a ventral chain of ganglia. There is a median dorsal eye. Cyclopoids are intermediate hosts of the parasitic guinea worm (*Dracunculus medinensis*), and sometimes the fish tapeworm

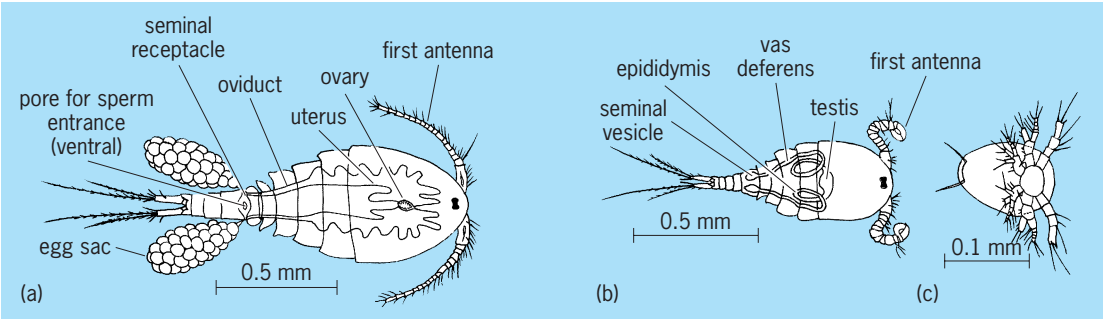

Fig. 1. *Cyclops vernalis*. (*a*) Dorsal view of female. (*b*) Dorsal view of male. (*c*) Ventral view of the first of the six nauplius stages.

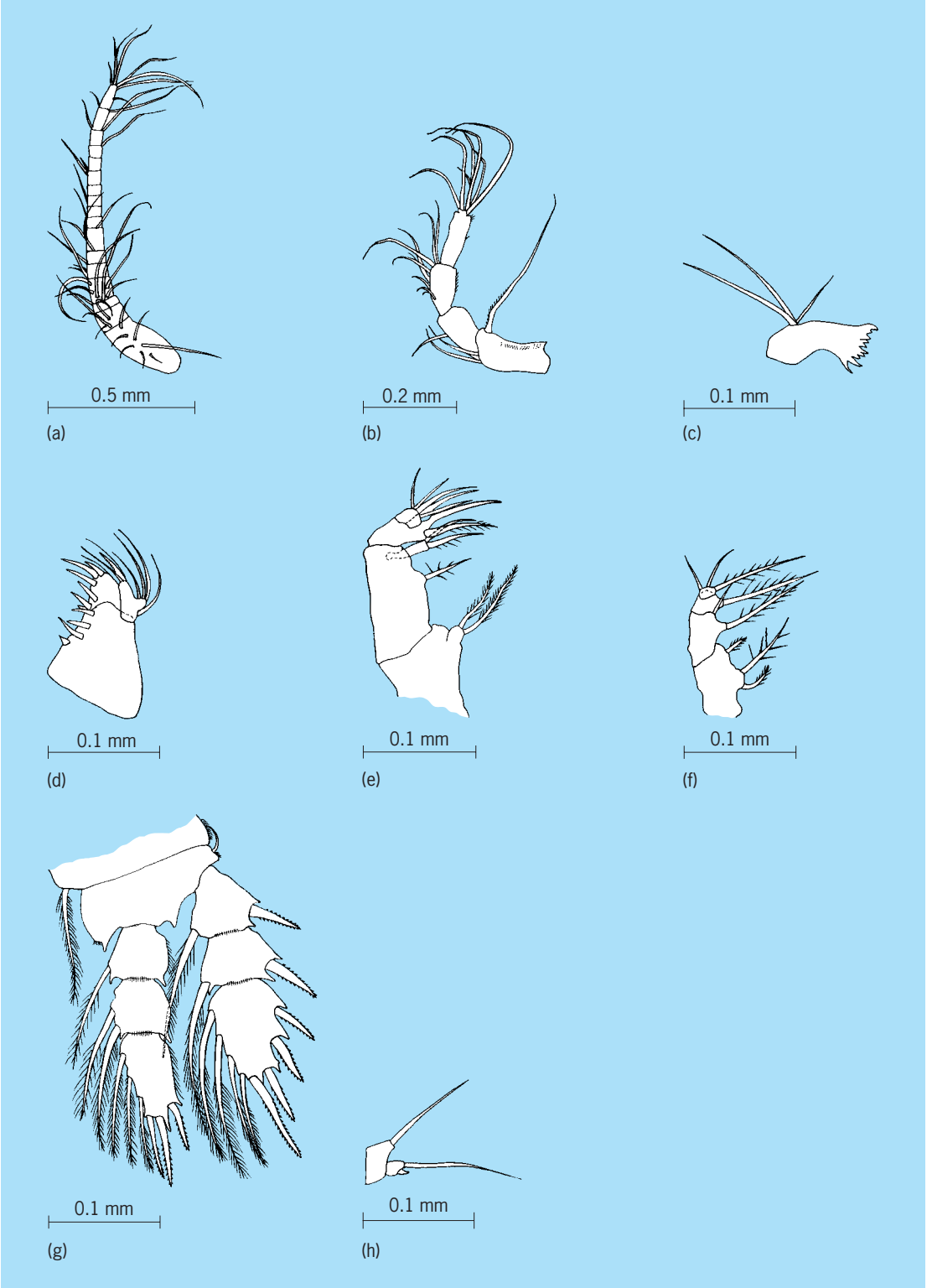

Fig. 2. Representative appendages of *Cyclops*. (*a*) First antenna of female. (*b*) Second antenna. (*c*) Mandible. (*d*) First maxilla. (*e*) Second maxilla. (*f*) Maxilliped. (*g*) Typical swimming leg (right third leg). (*h*) Fifth leg.

(*Dibothriocephalus latus*). Most fresh-water species live in shallow water, swimming from plant to plant, but salt-water species are generally water-treaders. Food is not secured by filtration, but is seized and eaten with the biting mouthparts. Many species have a worldwide distribution. *See* COPEPODA.

Harry C. Yeatman

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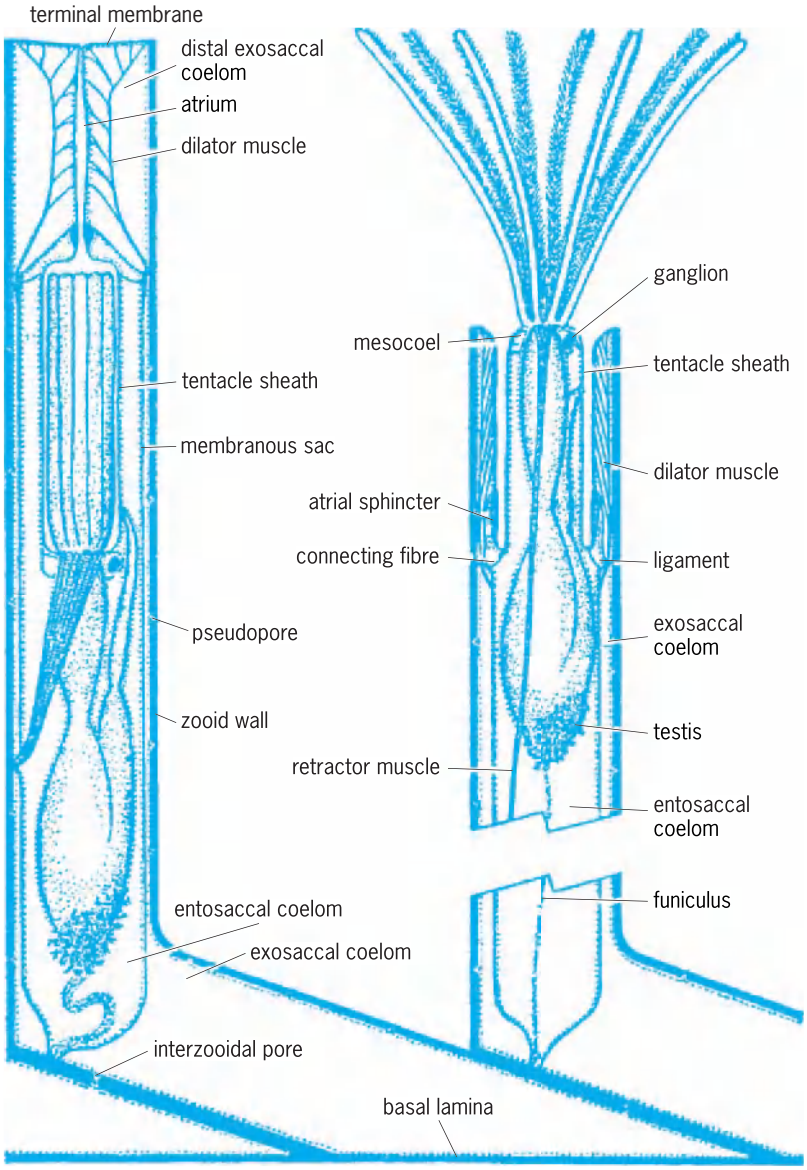
Cyclostomata (Bryozoa)

An order of bryozoans in the class Stenolaemata. Cyclostomes tend to have delicate colonies composed of relatively isolated, loosely bundled or tightly packed, comparatively simple, long, slender, and tubular zooecia (individual living chambers), with thin, highly porous, calcareous walls. When regionated, their colonies display a very gradual transition between rather indistinct endozones and exozones. *See* BRYOZOA; STENOLAEMATA.

Morphology. Cyclostome colonies commonly are small and delicate, but some are moderately large and massive. They may be inconspicuous encrusting threadlike networks (*Corynotrypa*, *Hederella*), delicate prostrate branches (*Tubulipora*, *Diaperoecia*), thin encrusting disks (*Lichenopora*) or sheets (*Berenicea*), nodular masses (*Ceriopora*), or erect tuftlike (*Crisia*), twiglike (*Tretocycloecia*), or frondlike (*Hornera*) growths. Cyclostome colonies may sometimes bear rhizoids (rootlike structures that reinforce the colony's attachment to a substrate), nanozooecia (dwarf zooecia), mesopores (tiny pores), and various small coelomic chambers, but lack many of the specialized morphologic structures characteristic of other bryozoan orders. Many colonies also display well-developed ovicells (brood chambers) that have evolved by complete modification of initially normal-appearing zooids (gonozooids).

Individual cyclostome zooecia are straight to slightly curved long tubes, each containing a polyp with tentacles (see **illustration**). The walls of adjacent zooecia may be either distinctly separate or fused together. The zooecial walls, generally thin throughout the colony and having finely granular or finely laminated microstructure, consist of calcareous material (calcite) perforated by many tiny pores which are either open (interzooidal or communication pores) or filled (in living forms) with soft tissues (pseudopores). In addition, some species may possess a chitinous cuticle covering the external surface of the outermost zooecial walls in the colony. The space within each zooecium usually lacks transverse calcareous partitions, but sometimes is crossed by a few diaphragms. Round or circular in outline, the zooecial aperture lacks an operculum, but in some forms is surrounded by a raised rim (peristome). Cyclostome zooids lack epistomes, have few to moderate numbers of tentacles, and (like the colony overall) can be either hermaphroditic or dioecious.

Life cycle. Immediately after fertilization, the cyclostome zygote enters an ovicell and divides into as many as 100 embryos (polyembryony), each of which develops into a larva. After escaping from the ovicell, the larva swims freely for less than a day before settling down to the bottom and attaching to a substrate. There, the larva metamorphoses (by means of complete degeneration of its internal structures) into an ancestrular (primary) zooid from which repeated asexual budding produces the rest of the colony (but no resistant resting bodies).



Cyclostome zooids with tentacles retracted versus extended. (*Reprinted with permission from R. J. Cuffey and J. E. Utgaard, Bryozoans, in R. Singer, ed., Encyclopedia of Paleontology, Fitzroy Dearborn Publishers, 1999*)

History and classification. Exclusively marine, cyclostomes (recently termed Tubuliporata by some) are known first from the Early Ordovician, when they may have evolved from ctenostomes, and possibly gave rise to the other Paleozoic stenolaemate orders. Rare and inconspicuous throughout the rest of the Paleozoic and Triassic, cyclostomes became moderately common in Jurassic and Cretaceous seas, but have since (including today) been less numerous. However, cyclostomes are of great scientific interest because, as the only living stenolaemates, they can shed much light on related extinct forms. *See* CTENOSTOMATA.

Roger J. Cuffey

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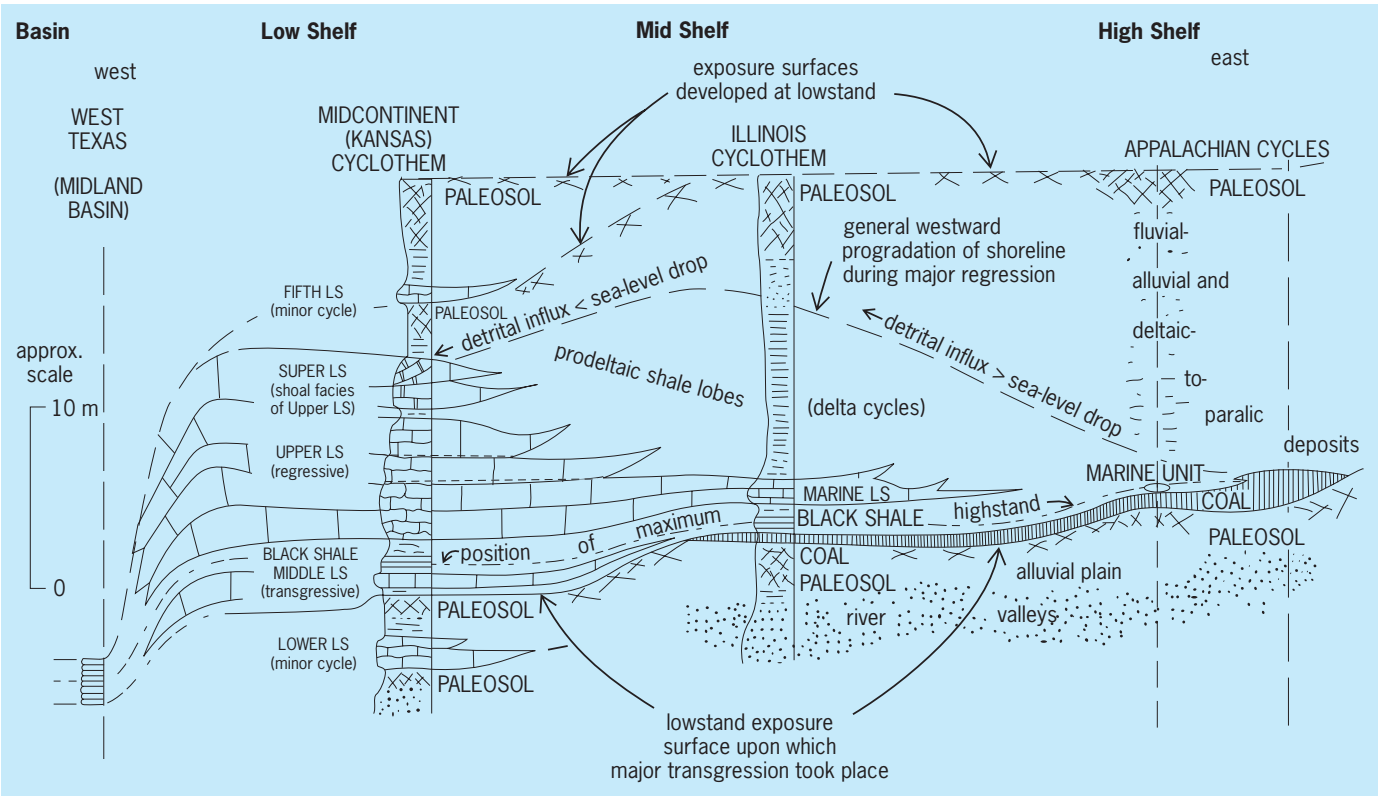
Cyclothem

A vertical sequence of several different kinds of distinctive sedimentary rock units that is repeated upward through the stratigraphic succession. Originally defined in the rock succession of Pennsylvanian age in the Illinois Basin in the 1930s, the rock types include coal, limestone, sandstone, and several types of shale and mudstone. Cyclothems were soon recognized elsewhere in rocks of this age in the central and eastern United States. Those in the Midcontinent (Kansas and states to the northeast) are dominated by several types of limestone and shale, with less coal and sandstone. Those in the Appalachian region are dominated by coal, mudstone, shale, and sandstone, with less limestone. Some workers generalized the term, and simple cyclothems became recognized in coal-bearing successions of other ages, but the term never achieved wide usage outside of rocks of Pennsylvanian age. *See* PENNSYLVANIAN; SEDIMENTARY ROCKS; STRATIGRAPHY.

The original definition stated that this sequence of rocks was deposited during a single sedimentary cycle of the type that prevailed during the Pennsylvanian Period, thus implying a strong genetic relation among such sequences in strata of this particular age. Three major types of origin were proposed: (1) periodic uplift (tectonism) of the active Appalachian Mountains followed by subsidence and encroachment of the sea; (2) periodic rise and fall of sea level (eustacy) driven by repeated episodes of continental glaciation in the southern continents of that time; and (3) shifting of deltas when flooding broke the levee and established a new delta in a different but overlapping area. The most recent work on Pennsylvanian cyclothems (when complex cyclothems were best developed) has resulted in the recognition of well-developed widespread paleosols (ancient soils) beneath the coals nearly everywhere and the widespread lateral continuity of the marine limestones and shales in the Midcontinent. This, in conjunction with the absence of deltas in most of that region, has ruled out delta-shifting as the basic cause. The estimates of rather great water depth attained during marine inundation in the Midcontinent (which was lower on the ancient shelf) along with estimates of greater frequency than is



Restored cross section showing relations of cyclothems in Midcontinent, Illinois, and Appalachian regions. LS = Low Shelf. (*After P. H. Heckel, Evaluation of evidence for glacio-eustatic control over marine Pennsylvanian cyclothems in North America and consideration of possible tectonic effects, in Tectonic and Eustatic Controls on Sedimentary Cycles, ed. by J. M. Dennison and F. R. Ettensohn, SEPM Concepts in Sedimentology and Paleontology, 4:65–87, 1994; and P. H. Heckel, Glacial-eustatic base-level–climatic model for late Middle to Late Pennsylvanian coal-bed formation in the Appalachian Basin, J. Sed. Res., B65:348–356, 1995*)

likely with tectonic fluctuations of sea level have ruled out tectonism as a general cause. Estimates of cyclothem frequency and magnitude of water-depth changes are both within the ranges detected for the well-known Pleistocene continental glaciations, and thus further support the glacial eustatic origin for cyclothems. For this reason (combined with the specific original definition), the term is probably most usefully applied only to glacially driven repeating sequences of rock units. This would include strata from latest Mississippian through Pennsylvanian to mid-Permian age (most of late Paleozoic time), and the term could apply to similar genetic sequences of rock in glacially influenced late Cenozoic time, which includes the Pleistocene ice ages and modern times. *See* CENOZOIC; MISSISSIPPIAN; PALEOSOL; PERMIAN; PLEISTOCENE.

The sequence of events that explains Pennsylvanian cyclothems of the central and eastern United States is shown in the **illustration**. At sea-level lowstand, sandstones were deposited in ancient valleys while soils formed over the exposed land surface on the areas between the rivers. As the glaciers melted, sea level rose causing coal swamps to form in poorly drained areas landward of the encroaching sea. These were followed by shallow-water marine limestone, which in the Midcontinent was followed by deep-water black phosphatic shale deposited when the sediment surface passed below the photic zone in this deeper area, and sea-circulation changes diminished bottom oxygen and concentrated phosphate. In the shallower Appalachian region, the marine limestone was the deepest water deposit. Sea-level fall then caused outbuilding of deltas in the Appalachian region, followed by subaerial exposure and development of paleosols, but caused formation of shoaling-upward limestone above the black shale in the Midcontinent, an area that was not affected by delta outbuilding or subaerial exposure until later.

Cyclothems were first noticed in the United States because of the many distinctive rock types involved. These resulted from the interplay between the then tropical to equatorial humid climate with ready access to detrital sand and mud and the marine circulation changes caused by the changing water depth of the sea. Coal-rich cyclothems were soon recognized in strata of the same age in western Europe. Currently, cyclothems without coal are being identified in late Paleozoic strata elsewhere in the world by the recognition of thin paleosols and exposure surfaces in more monotonous limestone successions in the more subtropical regions of that time, where the climate was drier and lithic differences not as conspicuous. Complex cyclothems are also being recognized in repeating sequences of entirely subtidal marine rock units in late Paleozoic pelagic realms. *See* PALEOZOIC.

Philip H. Heckel

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Cyclotron resonance experiments

The measurement of charge-to-mass ratios of electrically charged particles from the frequency of their helical motion in a magnetic field. Such experiments are particularly useful in the case of conducting crystals, such as semiconductors and metals, in which the motions of electrons and holes are strongly influenced by the periodic potential of the lattice through which they move. Under such circumstances the electrical carriers often have "effective masses" which differ greatly from the mass in free space; the effective mass is often different for motion in different directions in the crystal. Cyclotron resonance is also observed in gaseous plasma discharges and is the basis for a class of particle accelerators. *See* BAND THEORY OF SOLIDS; PARTICLE ACCELERATOR.

The experiment is typically performed by placing the conducting medium in a uniform magnetic field H and irradiating it with electromagnetic waves of frequency ν. Selective absorption of the radiation is observed when the resonance condition $\nu = qH/2\pi m^*c$ is fulfilled, that is, when the radiation frequency equals the orbital frequency of motion of the particles of charge q and effective mass m^* (c is the velocity of light). The absorption results from the acceleration of the orbital motion by means of the electric field of the radiation. If circularly polarized radiation is used, the sign of the electric charge may be determined, a point of interest in crystals in which conduction may occur by means of either negatively charged electrons or positively charged holes. *See* HOLE STATES IN SOLIDS.

For the resonance to be well defined, it is necessary that the mobile carriers complete at least $^1/_2\pi$ cycle of their cyclotron motion before being scattered from impurities or thermal vibrations of the crystal lattice. In practice, the experiment is usually performed in magnetic fields of 1000 to 100,000 oersteds (1 oersted = 79.6 amperes per meter) in order to make the cyclotron motion quite rapid ($\nu \approx$ 10–100 gigahertz, that is, microwave and millimeter-wave ranges). Nevertheless, crystals with impurity concentrations of a few parts per million or less are required and must be observed at temperatures as low as 1 K in order to detect sharp and intense cyclotron resonances.

The resonance process manifests itself rather differently in semiconductors than in metals. Pure, very cold semiconductors have very few charge carriers; thus the microwave radiation penetrates the sample uniformly. The mobile charges are thus exposed to

radiation over their entire orbits, and the resonance is a simple symmetrical absorption peak.

In metals, however, the very high density of conduction electrons present at all temperatures prevents penetration of the electromagnetic energy except for a thin surface region, the skin depth, where intense shielding currents flow. Cyclotron resonance is then observed most readily when the magnetic field is accurately parallel to the flat surface of the metal. Those conduction electrons (or holes) whose orbits pass through the skin depth without colliding with the surface receive a succession of pulsed excitations, like those produced in a particle accelerator. Under these circumstances cyclotron resonance consists of a series of resonances $n\nu = qH/2\pi m^*c$ ($n = 1, 2, 3, \ldots$) whose actual shapes may be quite complicated. The resonance can, however, also be observed with the magnetic field normal to the metal surface; it is in this geometry that circularly polarized exciting radiation can be applied to charge carriers even in a metal.

Cyclotron resonance is most easily understood as the response of an individual charged particle; but, in practice, the phenomenon involves excitation of large numbers of such particles. Their net response to the electromagnetic radiation may significantly affect the overall dielectric behavior of the material in which they move. Thus, a variety of new wave propagation mechanisms may be observed which are associated with the cyclotron motion, in which electromagnetic energy is carried through the solid by the spiraling carriers. These collective excitations are generally referred to as plasma waves. In general, for a fixed input frequency, the plasma waves are observed to travel through the conducting solid at magnetic fields higher than those required for cyclotron resonance. The most easily observed of these excitations is a circularly polarized wave, known as a helicon, which travels along the magnetic field lines. It has an analog in the ionospheric plasma, known as the whistler mode and frequently detected as radio interference. There is, in fact, a fairly complete correspondence between the resonances and waves observed in conducting solids and in gaseous plasmas. Cyclotron resonance is more easily observed in such low-density systems since collisions are much less frequent there than in solids. In such systems the resonance process offers a means of transferring large amounts of energy to the mobile ions, a necessary condition if nuclear fusion reactions are to occur. *See* NUCLEAR FUSION; PLASMA (PHYSICS). Walter M. Walsh

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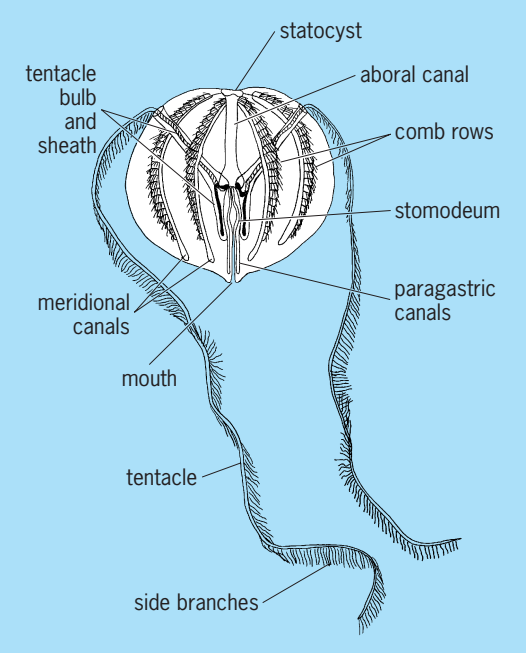
Cydippida

The largest order of the phylum Ctenophora, comprising 5 families (Bathyctenidae, Haeckelidae, Lampeidae, Mertensidae, and Pleurobrachiidae) and 11 genera. The body of cydippids is usually globular or cylindrical in shape (see **illustration**); sizes range from a few millimeters to about 15 cm (6 in.). Most



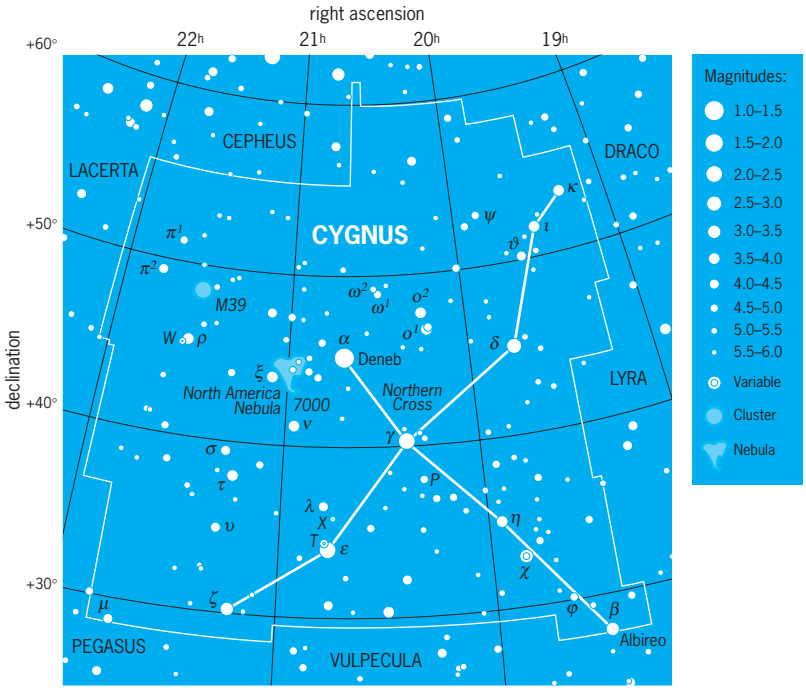
Typical *Pleurobrachia* sp.

cydippids are colorless and transparent, but some deep-sea species are pigmented dark red. The comb rows are of equal length, and the eight meridional canals lying beneath them end blindly at the oral end. Adult cydippids retain the larval morphology common to all ctenophores (except Beroida) and are usually thought to be the most primitive order. All forms have two main tentacles, which have side branches (tentilla) in some species. Cydippids catch prey on the outstretched tentacles, contracting them to bring it to the mouth. In three families (Bathyctenidae, Haeckelidae, and Lampeidae) the tentacles emerge from the body near the oral end; these cydippids have extensile mouths and catch large gelatinous animals. In the Pleurobrachiidae and Mertensidae, the tentacles exit near the aboral end, have numerous side branches, and catch mostly small crustacean prey. The well-known genus *Pleurobrachia* is widely distributed in temperate coastal waters and estuaries, where it can be a major predator of copepods and larval fish. Many other species of cydippids occur in oceanic waters, from the surface to mesopelagic depths. *See* BEROIDA; CTENOPHORA.

Laurence P. Madin

Cygnus

The Swan, a large constellation lying prominently along the northern Milky Way (see **illustration**). It contains the stars that define the asterism (star pattern) known as the Northern Cross. Near Deneb, its brightest star and the nineteenth brightest star in the sky, the North America Nebula shows its glowing gas and dark outlining dust. Deneb marks the swan's tail while Albireo, a double star whose components beautifully contrast in color, marks the swan's head.



Modern boundaries of the constellation Cygnus, the Swan. The celestial equator is 0° of declination, which corresponds to celestial latitude. Right ascension corresponds to celestial longitude, with each hour of right ascension representing 15° of arc. Apparent brightness of stars is shown with dot sizes to illustrate the magnitude scale, where the brightest stars in the sky are 0th magnitude or brighter and the faintest stars that can be seen with the unaided eye at a dark site are 6th magnitude. (*Wil Tirion*)

Deneb in Cygnus, Vega in Lyra, and Altair in Aquila make the "summer triangle." *See* LYRA.

In Greek mythology, Phaethon, the son of the sun god, was allowed to drive the chariot of the Sun across the sky, but lost control. Zeus saved Earth from disaster by hurling a thunderbolt at Phaethon, who fell into the Eridanus River. Eridanus is now another constellation. A devoted friend of Phaethon dived into the water to search for his friend. Zeus changed him into the swan, changed his name to Cygnus, and elevated him to the heavens.

The modern boundaries of the 88 constellations, including this one, were defined by the International Astronomical Union in 1928. *See* CONSTELLATION.

Jay M. Pasachoff

Cylinder

The solid of revolution obtained by revolving a rectangle about a side is called a cylinder, or more precisely a right circular cylinder. More generally the word cylinder is used in solid geometry to describe any solid bounded by two parallel planes and a portion of a closed cylindrical surface. In analytic geometry, however, the word cylinder refers not to a solid but to a cylindrical surface (see **illus.**). This is a surface generated by a straight line which moves so that it always intersects a given plane curve called the directrix, and remains parallel to a fixed line that intersects the plane of the directrix. (Part or all of the directrix may consist of straight-line segments.) The generating line in each of its positions is called an element of the cylinder, and a section of the cylinder made by a plane perpendicular to the elements is called a right section. All right sections of a cylinder are congruent. A circular cylinder (cylindrical surface) is one whose right sections are circles. Its other sections made by planes cutting all the elements are ellipses. Cylinders whose right sections are ellipses, parabolas, or hyperbolas are called elliptic cylinders, parabolic cylinders, or hyperbolic cylinders, respectively. All these are quadric cylinders. The lateral surface of a prism is a special type of cylindrical surface whose directrix and right sections are polygons.

Diagram of cylindrical surface.

The volume of a solid cylinder is $V = Bh$, where B denotes the area of the base, and h its altitude (height) measured perpendicular to the base. For a right circular cylinder whose base is a circle of radius r, the volume is $V = \pi r^2 h$, and the total surface area is $S = 2\pi r^2 + 2\pi rh$. *See* EUCLIDEAN GEOMETRY.

J. Sutherland Frame

Cyperales

An order of flowering plants, division Magnoliophyta (Angiospermae), in the subclass Commelinidae of the class Liliopsida (monocotyledons). The names Glumiflorae, Graminales, and Poales have also been used for this order. There are only two families, the Poaceae (Gramineae), with about 8000 species, and the Cyperaceae, with nearly 4000. The Cyperales are Commelinidae with reduced, mostly wind-pollinated or self-pollinated flowers that have a unilocular, two- or three-carpellate ovary bearing a single ovule. The flowers are arranged in characteristic spikes or spikelets representing reduced inflorescences. The perianth is represented only by a set of bristles or tiny scales, or is completely missing. The leaves generally have a well-defined sheath and a narrow blade, often with a small adaxial appendage (the ligule) at the junction of the two. The stomates have two supporting cells and are arranged in straight files or rows of one or two, all oriented in the same direction. The pollen is uniformly trinucleate, and vessels are present in all vegetative organs.

The Cyperaceae, or sedge family, includes the bulrushes (*Scirpus*) and the papyrus (*Cyperus papyrus*) of Egypt, as well as the sedges (*Carex*). The Poaceae, or grass family, embraces all true grasses, including bamboos and the cereal grains such as wheat, maize, oats, and rye. The two families differ in a number of more or less consistent technical characters of the inflorescence, fruits, stems, and leaves. The well-developed intercalary meristem at the base of the blade in the grass leaf permits the plants to withstand grazing, and may be largely responsible for the dominance of grasses in large areas of the world with a suitable climate (seasonal rainfall, with the dry season too severe to permit forest growth). *See* BAMBOO; BARLEY; CEREAL; COMMELINIDAE; CORN; FLOWER; GRASS CROPS; LILIOPSIDA; MAGNOLIOPHYTA; MILLET; OATS; PLANT KINGDOM; RICE; RYE; SORGHUM; SUGARCANE; TIMOTHY; WHEAT.

Arthur Cronquist; T. M. Barkley

Cypress

The true cypress (*Cupressus*), which is very close botanically to the cedars (*Chamaecyparis*). The principal differences are that in *Cupressus* the cones are generally larger and the cone scales bear many seeds, whereas in *Chamaecyparis* each cone scale bears only two seeds; and in *Cupressus* the branchlets are not flattened as in *Chamaecyparis*. *See* CEDAR; PINALES.

All *Cupressus* species in the United States are western and are found from Oregon to Mexico. The Arizona cypress (*Cupressus arizonica*) found in the southwestern United States and the Monterey cypress (*Cupressus macrocarpa*) of California are medium-sized trees and are chiefly of ornamental value. The Italian cypress (*Cupressus sempervirens*) and its varieties are handsome ornamentals, but usually do well only in the southern parts of the United States. Other trees are also called cypress, such as the Port Orford cedar (*Chamaecyparis lawsoniana*) known also as the Lawson cypress, and the Alaska cedar (*Chamaecyparis nootkatensis*), known also as the Nootka cypress or cedar.

The bald cypress (*Taxodium distichum*) is an entirely different tree that grows 120 ft (36 m) tall and 3–5 ft (1–1.5 m) in diameter. It is found in the swamps of the South Atlantic and Gulf coastal plains and in the lower Mississippi Valley. The soft needlelike leaves and short branches are deciduous; hence, they drop off in winter and give the tree its common name. Also known as the southern or red cypress, this tree yields a valuable decay-resistant wood used principally for building construction, especially for exposed parts or where a high degree of resistance to decay is required as in ships, boats, greenhouses, railway cars, and railroad ties. *See* FOREST AND FORESTRY; TREE.

Arthur H. Graves; Kenneth P. Davis

Cypriniformes

An order of actinopterygian fishes in the superorder Ostariophysi with about 2700 species, which ranks second in number only to the Perciformes. Cypriniformes is the largest order of freshwater fishes in the world. Six families of cypriniforms occupy the freshwaters of most of the major landmasses of the world, but are notably absent from South America, Madagascar, and the Australian realm. North America is represented by two native families, Cyprinidae and Catostomidae.

Cypriniforms differ from other fishes in having the following combination of characters: The fifth ceratobranchial (a pharyngeal bone) is enlarged and teeth (pharyngeal teeth) are ankylosed, or fused, to the bone; the pharyngeal teeth are opposed to a pair of enlarged posterior processes of the basioccipital bone; the mouth (jaws and palate) is toothless; the premaxillae have an ascending process and are usually protractile; four vertebrae are in the Weberian apparatus (a series of modified vertebrae that form a chain connecting the swim bladder with the inner ear); there are three branchiostegal rays; an adipose fin is lacking (except in some loaches); and, with a few exceptions, the head is scaleless. *See* OSTARIOPHYSI.

Cyprinidae. Shiner, minnow, chub, carp, and dace are the most common names applied to North American cyprinids. Rudd, ide, tench, and especially carp are cyprinid common names heard in Eurasia (see **illustration**).

With about 210 genera and more than 2000 species (and counting), this is the largest freshwater family of fishes, indeed, with the possible exception of the Gobiidae, the largest of all the vertebrate families. About 300 native species in 52 genera, in addition to seven genera and eight established nonnative species, occur in North America. Seven subfamilies are recognized but only one, the

subfamily Leuciscine, represents the native North America cyprinids.

Cyprinids are known from the Eocene in Asia and from the Oligocene in Europe and North America. The majority of the species, and morphologically the most diverse, are in Eurasia, especially in China and southeastern Asia.

Morphology. Cyprinids are distinguished from other cypriniforms in the following ways: pharyngeal teeth are in one to three rows, with never more than eight in any row; the upper jaw is bordered by premaxillae and is usually protrusible; the lips are usually thin, and neither plicate (pleated) or papillose; barbels are either present or absent; the first dorsal fin is spikelike in some species; and intermuscular bones are present. Most cyprinids are small fishes; the smallest, *Danionella transluida*, a species from Burma, is not known to exceed 12 mm (less than $^1/_2$ in.). The largest species are *Catlocarpio siamensis* of Thailand, whose length is no less than 2.5 m (probably to 3 m), and *Tor putitora* of eastern India, which is about 2.7 m. The largest native North American minnows are *Ptychocheilus lucius* (Colorado pikeminnow), which is reported to have reached 1.8 m (5.9 ft) and *P. oregonensis* (northern pikeminnow), 1.22 m (4 ft). However, most cyprinids of the world are probably less than 25 cm in length.

Reproduction. Reproductive strategies vary greatly. The large carps typically spawn in large free-flowing rivers where they broadcast millions of pelagic eggs which develop while carried by the river current. The eggs hatch within a few days and the fry reach flooded backwaters where they mature. The smaller species produce far fewer eggs, but compensate by preparing a nest in the form of a depression excavated in the substrate, or by selecting protective cover, such as crevices in rocks or bark of fallen trees, for secure sites to deposit their fertilized eggs. This strategy calls for parental care, usually provided solely by the male parent. Some species deposit their fertilized eggs in the interstices of gravel and pebbles, and no parental care is provided.

Many cyprinids are silvery and without prominent color markings, but during the breeding season some species acquire conspicuous red, yellow, and blue colors, which may be accompanied by iridescences. Also, during the breeding season males, and some females, develop tubercles on the head, body, and fins. Depending on size, number, and placement, the tubercles may function in defense, nest building, or to hold breeding pairs close together during spawning.

It is not unusual for a female minnow to produce in a single reproductive season a mass of mature eggs that exceeds several times her body weight. Such fecundity is accomplished by producing eggs in multiple clutches, the clutches being spaced to take advantage of environmental conditions likely to promise a successful spawn.

Yet another reproductive strategy is seen in the bitterling (*Rhodeus*) of Central Europe, which ensures the safety of its offspring by depositing the eggs within the valves of mussels. When the female

Koi carp (*Cyprinus carpio carpio*). (*Photo © by J. J Photo*)

is ready to spawn the oviduct is drawn out as a long tube that acts as an ovipositor, and the eggs are fertilized after they are given over to the mussel. The respiratory currents of the mussel provide essential aeration necessary for development of the eggs. The mussel's service is rewarded—its own embryos attach to the gill of the bitterling, where, safe from predators, they undergo early stages of development. *See* CARP.

Psilorhynchidae. Members of this small family (two genera and five species), known from the mountain streams of Nepal and adjacent India to western Burma, have a small subterminal mouth whose jaws have sharp horny edges and fleshy lips; a pharyngeal bone with one row of four teeth; no barbels; narrow gill openings; a flattened ventral head surface; and reduced swim bladder.

Gyrinocheilidae. This family, known as algae eaters, is from Southeast Asia. The single genus (*Gyrinocheilus*), with four species, has several distinguishing characters which adapt it to an existence in high gradient streams. Rare among cypriniforms is the absence of pharyngeal teeth. The mouth is modified as a sucking device for attaching to objects; the gill opening is reduced to two small apertures, an upper incurrent and lower excurrent; and gill rakers number about 140. These fishes feed exclusively on algae, hence are popular as glass cleaners for home aquaria.

Catostomidae. The suckers are North American and eastern Asian fishes comprising 13 genera and about 72 species, two of which are thought to be extinct. North America has the greatest diversity and would appear to be the center of origin of catostomids. However, the evidence seems to indicate that catostomids originated in Asia from cyprinid ancestry.

Catostomids are characterized by having a single row of pharyngeal teeth numbering 16 or more. In some species the pharyngeal teeth are numerous, slender, very closely set (baleenlike), and function as a sieve; in other species the teeth are more or less comblike and function for mastication, or molarlike and used for grinding. The jaws lack teeth, the lips are thick and plicate or papillose, and both the premaxilla and maxilla border the upper jaw. Barbels are absent and scales are absent from the head.

The family has three rather distinctive subfamilies.

The Ictiobinae, with two genera, *Carpiodes* (quillbacks and carpsuckers) and *Ictiobus* (buffalo suckers), have a deep, relatively strongly compressed body; large scales; a complete lateral line; 115 to 190 pharyngeal teeth on each pharyngeal arch; a long dorsal fin base; and a swim bladder with two chambers. They occur from Canada to Guatemala. The Cycleptinae, having an elongate body in the adults, very small head, dorsal fin falcate (deeply concave) and with a long base, and an air bladder with two lobes, is represented in China by one species, *Myxocyprinus asiaticus*, and in North America (Mississippi River and adjacent gulf coastal drainages) by two described species of *Cycleptus*. The Catostominae have a slender terete (nearly cylindrical) body, small head, short dorsal fin, and swim bladder in two or three chambers. The lateral line is complete, incomplete, or absent, depending on the genus. Two genera, *Catostomus* (tribe Catostomini) and *Moxostoma* (tribe Moxostomatini), with about equal number of species, account for 42 of the 62 species of the family. Members of the tribe Catostomini live mostly in western North America, however one, *Catostomus catostomus*, ranges from New York State to eastern Siberia, and another, *C. commersoni* (white sucker), ranges from Georgia to British Columbia. All but one of the 30 species of Moxostomatini live in North America, east of the Continental Divide.

Most catostomids feed on the bottom on a variety of foods. Depending on the species, the food may be algae, detritus, aquatic insect larvae, or mollusks. Typically catostomids make spring runs up small streams, where they select clean gravel in flowing water as a spawning site. No parental care is given the young, except among the chubsuckers (genus *Erimyzon*), in which the males defend a nest, often an unoccupied nest of the stoneroller or creek chub.

Cobitidae. The loaches consist of about 110 species in 18 genera in the freshwaters of Eurasia plus one genus in Morocco. The greatest diversity is in southern Asia. This family and the river loaches are the most atypical of the cypriniforms. Cobitids are identified by their eel-like to fusiform shape, a subterminal mouth, pharyngeal teeth in one row, and an erectile spine below the eye. Two subfamilies are recognized: the Corbitinae, with one pair of rostral barbels and caudal fin usually rounded or slightly emarginate, and the Botiinae, with two pairs of rostral barbels and caudal fin deeply forked. Loaches are bottom-dwelling fishes, with a maximum length about 40 cm. A number of species are popular aquarium fishes.

Balitoridae. This family (river loaches) are similar to the cobitids but are recognized as a different family primarily by the Weberian apparatus, which has a Y-shaped tripus (the posteriormost Weberian ossicle). Also, the balitorids have three or more pairs of barbels near the mouth. Two rather distinctive subfamilies are recognized. The Nemacheilinae contains about 350 species of Eurasian fishes, with the greatest concentrations of species in India, Indochina, and China. They lack a spine under the eye; they have two pairs of rostral barbels and one pair of maxillary barbels, a subterminal mouth, and an adiposelike fin in some species; and scales are present or absent. Distinguishing characters of the subfamily Balitorinae are paired fins that are enlarged with adhesive pads on the ventral surface and oriented horizontally; pelvic fins, in some species, that are united under the belly; and a flattened head and body. These adaptations fit them well for occupying the fast mountain streams from India throughout Southeast Asia.

Herbert Boschung

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Cyprinodontiformes

An order of teleost fishes comprising eight families, 88 genera, and about 810 species that inhabit fresh and brackish waters of tropical and temperate North and Middle America, Eurasia, and Indo-Malaysia. Several features unify the order: a symmetrical caudal skeleton of one epural supporting an unlobed (round) caudal fin; low-set pectoral fins with a large, scalelike postcleithrum; sensory pores chiefly on the head; and the upper jaw bordered by protractile premaxillae. Sexual dimorphism is common, with the males brightly colored, and, in the livebearers, an anal fin that is modified as a gonopodium (intromittent organ) for internal fertilization. The top of the head is usually flat and the mouth is usually terminal to superior, features that adapt the fishes to feed on or near the surface of the water.

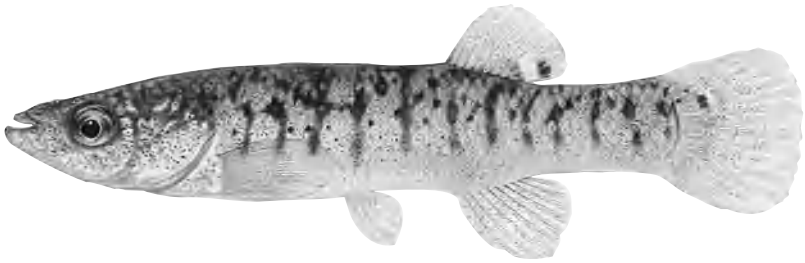
Aplocheilidae. These fishes, known as rivulines, differ from all other cyprinodontiforms by having the pelvic fin bases inserted close together (versus pelvic fins not close together), a metapterygoid bone (versus no metapterygoid), and three basibranchials (versus two basibranchials). Two subfamilies are recognized: the Aplocheilinae (Old World rivulines) of Africa (south of the Sahara Desert to South Africa) Madagascar, Seychelles, India, Sri Lanka, and Indo-Malaysian Archipelago to Java; and the Rivulinae (New World rivulines) of southern Florida and

Middle America to Uruguay. In the Aplocheilinae, comprising eight genera and about 185 species, the supracleithrum is fused to the posttemporal and the first postcleithrum is present, whereas in the Rivulinae, comprising 12 genera and no less than 125 species, the supracleithrum is not fused to the posttemporal and the first postcleithrum is absent. *Rivulus marmoratus* (mangrove rivulus) of southern Florida and the West Indies is a self-fertilizing hermaphrodite. Another oddity in this group is *R. nudiventris* of Brazil, which lacks pelvic fins and a pelvic girdle. Reproduction is oviparous, and in only a few species is fertilization internal.

Profundulidae. This family of one genus, *Profundus*, and five species occurs in the Atlantic and Pacific slopes of Mexico, Guatemala, and Honduras.

Goodeidae. Two subfamilies are recognized. The Goodeinae, comprising 43 species in 16 genera, are centered in the Rio Lerma River basin of Mexico. They are viviparous fishes, and internal fertilization is accomplished by a pseudophallus fashioned from crowded short anterior anal fin rays that are slightly separated by a notch from the rest of the fin. The eggs are small, with little yolk, but placenta-like structures provide the embryos with nutrients and oxygen. The four species in the subfamily Empetrichthyinae lack a pelvic girdle and pelvic fins, as well as a pseudophallus, and fertilization is external. They are further distinguished from the other goodeids by lacking a lateral line and by the placement of the dorsal and anal fins far back on the body. These fishes are limited to a few pools and springs in southern Nevada. Two Mexican species and one Nevada species, the Ash Meadows poolfish, are thought to be extinct.

Fundulidae. Topminnows and killifishes comprise five genera and 48 species. Forty species occur in North America, 35 of which are in the genus *Fundulus* (see **illustration**). Most of these fishes are elongate, and somewhat compressed and flattened about the head and nape. The mouth is small and turned upward; the jaws are protractile and the lower extends beyond the upper; the dorsal and anal fins are small, similar in size and shape and posteriorly placed nearly opposite each other; the anal fin is unmodified; and fertilization is external. The flat head and superior mouth allow them to feed at the surface, as well as to utilize the oxygen-rich layer at the atmosphere-water interface without greatly altering their usual swimming posture. Consequently, these fishes can survive in habitats where the subsurface waters are essentially depleted of oxygen. They inhabit the lowlands of North and Middle America from southern Canada to Yucatan, as well as Bermuda and Cuba. Some of the fundulids are very euryhaline and temperature tolerant. The diamond killifish (*Adinia xenica*) can tolerate temperatures as high as 38°C (100°F) in freshwater with a salinity of 54 parts per thousand. It is not unusual to find Gulf killifish (*Fundulus grandis*) in tidal pools where the salinity reaches 76 ppt, as well as in the freshwater Brazos River 400 km (249 mi) from the Gulf of Mexico.



Female marsh killifish (*Fundulus confluentus*). (*Illustration © by Joseph R. Tomelleri*)

Valenciidae. This is a family of one genus and two species in the freshwaters of Spain, Italy, and western Greece.

Cyprinodontidae. Known as the pupfishes, this is a family comprised of nine genera and about 107 species of small freshwater, brackish water, and coastal marine inhabitants of North and Middle America, the West Indies, parts of South America, northern Africa, and the Mediterranean Anatolian area. Cyprinodontids are deep bodied and usually have a deep, strongly compressed caudal peduncle, an upturned mouth, and cycloid scales. The anal fin is unmodified, fertilization of eggs is external, and reproduction is oviparous. The dorsal process of each maxillary is expanded laterally, and the two processes nearly meet in the midline forming a distinct groove.

Five genera and 44 species of pupfishes occur in North America, 36 of which are in the genus *Cyprinodon*. The majority inhabit springs, sinkholes, salt marshes, and saline creeks in the deserts of Mexico and the southwestern United States. Various species of pupfishes are known to survive in pools that reach temperatures of 45°C (113°F), have salinity of 142 ppt (open oceanic water is about 35 ppt), and oxygen concentrations as low as 0.13 mg/liter. No other fishes could survive these harsh conditions. Some species have very restricted ranges (even a single small spring), which makes them especially vulnerable to extinction because of environmental changes, such as lowering ground water. Indeed, four species are extinct, three in Mexico and one, the Santa Cruz pupfish, which was described as new in 2002.

Anablepidae. This is a family of three genera and eight species inhabiting freshwater and brackish water of southern Mexico to southern South America. Unlike other cyprinodontiforms, the insertion of the pelvic fins is well behind the tip of the pectoral fins. Two of the genera have a tubular gonopodium formed from the anal fin rays, and in some males the gonopodium can be moved only to the right, in others only to the left. In females the genital aperture is open only to the right or only to the left. This probably means that a sinistral male only mates with a dextral female and vice versa. Fertilization is internal and the young are born alive. In the genus *Anableps* the eyes bulge above the top of the head and are divided longitudinally into upper

and lower portions, each portion with a pupil; thus, with the water line in the center of the eye, the fish can focus on images simultaneously from above and below. There are three species of the four-eyed fishes in the lowlands of the Pacific slope of southern Mexico to Honduras and the Atlantic slope of northern South America.

Poeciliidae/Poeciliinae. The livebearing subfamily Poeciliinae is considered by some authors to be equivalent to the family Poeciliidae. It comprises 22 genera and 192 species ranging from the eastern United States south to the Rio de la Plata drainage in northern Argentina, with the greatest diversity in Central America, Mexico, and the West Indies. Twelve genera and 93 species occur in North America. The males have a gonopodium formed from three anal rays. To accommodate the elongate gonopodium, the anal fin is displaced forward, often to a point under the pelvic fins. Depending on the species, poeciliid embryos are nourished by yolk deposited before the eggs are fertilized (ovoviviparity), or the embryos are nourished by nutrients transferred from the mother via a placenta (viviparity). A phenomenon commonly seen in poeciliids is superfetation, a process by which two or more broods at different stages of embryonic development occur simultaneously in the same female. The Amazon molly, *Poecilia formosa*, from Texas and northern Mexico, is unisexual, all female. They are of hybrid origin and depend on males of bisexual species for sperm to stimulate embryonic development. Many of the poeciliids are popular aquarium fishes, such as sailfin mollies, guppies, swordtails, and platyfishes. The mosquitofishes, *Gambusia affinis* and *G. holbrooki*, have been introduced into numerous countries for the purpose of mosquito control; however, their presence is of questionable value. As a result of predation on eggs, larvae, and juveniles of other fishes, mosquitofishes often significantly reduce or eliminate native fishes. Herbert Boschung

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Cyrtosoma

A subphylum of the Mollusca comprising members of the classes Monoplacophora, Gastropoda, and Cephalopoda. The use by paleontologists of subphyla, including Cyrtosoma and Diasoma, in the classification of the phylum Mollusca has not been generally adopted by students of living forms. There is no disagreement over the further division of the phylum into eight clearly distinguishable classes.

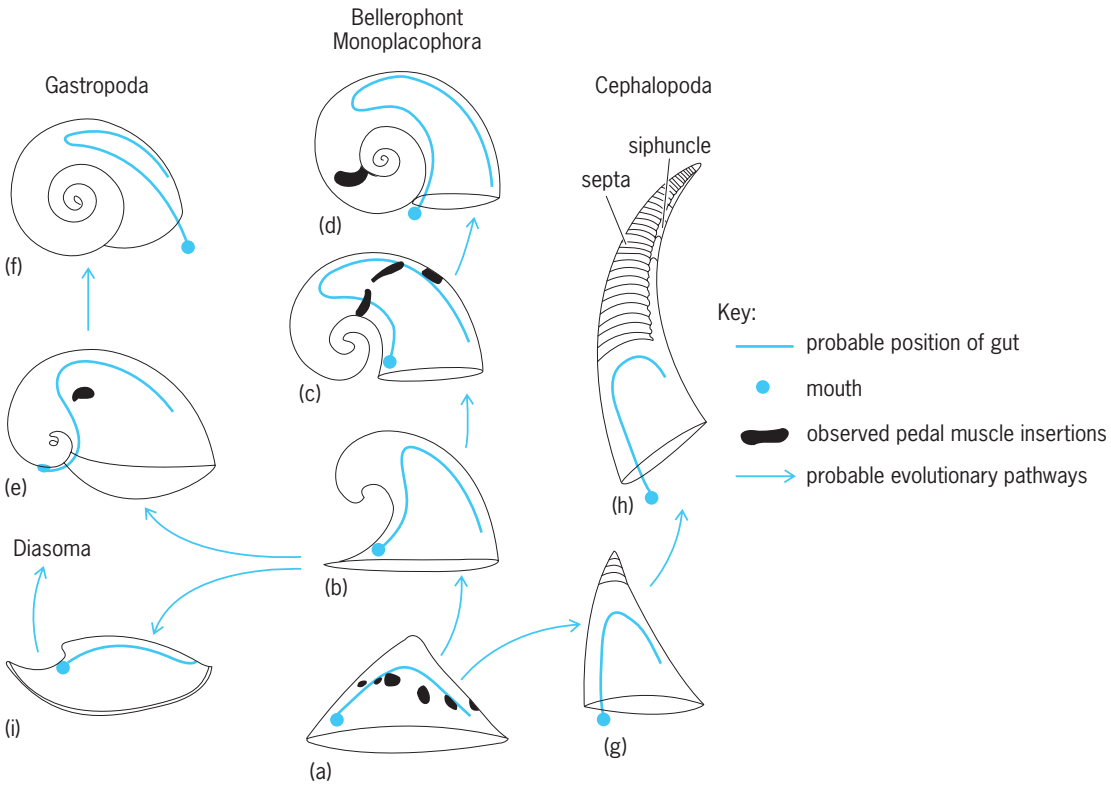
Primitive members of the Cyrtosoma have, or had, a conical univalved shell, often twisted into a spiral. The relatively small single-shell aperture forced the anus to lie close to the mouth, either behind it (Monoplacophora and Cephalopoda) or ahead of it in those forms which have undergone torsion (Gastropoda; **illus.** *f*). Studies of the comparative anatomy and history of living and fossil mollusks indicate that monoplacophorans are the most primitive cyrtosomes and that the other two classes were derived independently from the Monoplacophora (see illus.). *See* CEPHALOPODA; GASTROPODA; MONOPLACOPHORA.

Although both gastropods and monoplacophorans are found in earliest Cambrian rocks in Siberia and China, intermediate forms which link the two classes (*Pelagiella*; illus. *e*) survived through the Cambrian. Apparently asymmetric coiling of the shell of some early monoplacophorans caused the body to begin to rotate with respect to the shell, thus initiating torsion. When this process was completed, the shell was carried so that it coiled posteriorly, and the first gastropods had evolved. Other early monoplacophorans, which remained bilaterally symmetrical, gave rise to animals with a variety of symmetrical shell forms, including those of a diverse extinct group, the Bellerophontida. The bellerophonts have usually been placed in the Gastropoda, but there is mounting evidence that they were untorted and hence monoplacophorans.

The third cyrtosome class, the Cephalopoda, did not appear until the end of the Cambrian. Its first members (illus. *b*) had small conical shells with numerous calcareous partitions (septa) at the apical end. Similar septate monoplacophorans are known from slightly older rocks, but these lacked a siphuncle, used for buoyancy control. It was the appearance of the siphuncle which enabled the first cephalopods to swim, and it has been responsible for the great post-Cambrian success of that class.

Each cyrtosome class has, in turn, been the most diverse molluscan group. In the Ordovician the cephalopods succeeded the monoplacophorans as the most diverse molluscan class, and they held this position until the close of the Mesozoic. They were followed by the gastropods, which by then had diversified into a great variety of aquatic and terrestrial habitats. It has been estimated that about 80% of the species of living mollusks are gastropods.

Although primitive cyrtosomes, including the only living shelled cephalopod, *Nautilus*, are shell-bearing animals, many advanced cyrtosomes have either lost or greatly reduced their shells (examples include slugs, nudibranchs, squids, and octopuses). One rare family of gastropods has even evolved a truly bivalved shell. Consequently, it is not any



Evolution of the Cyrtosoma. (*a*) *Lenaella*. (*b*) *Latouchella*. (*c*) *Cyrtolites*. (*d*) *Bucania*. (*e*) *Pelagiella*. (*f*) *Aldanella*. (*g*) *Knightoconus*. (*h*) *Plectronoceras*. (*i*) *Mellopegma*. Organisms *a–e*, *g*, and *i* are monoplacophorans, *f* is an early gastropod, and *h* is the oldest cephalopod.

particular shell form which unites all members of the Cyrtosoma; rather, it is a knowledge of their evolutionary history. *See* MOLLUSCA.

Bruce Runnegar

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Cystoporata

An extinct order of bryozoans in the class Stenolaemata. Cystoporates tend to have robust colonies composed of relatively simple, long, slender, and tubular zooecia (with thin, somewhat porous calcareous walls), separated by blisterlike vesicles (cystopores) stacked upon one another. Their colonies display a very gradual transition between comparatively indistinct endozone and exozone regions. *See* BRYOZOA; STENOLAEMATA.

Morphology. Cystoporate colonies range from small and delicate to large and massive; they can be thin encrusting sheets, tabular or nodular masses, or erect twiglike or frondlike growths. Some cystoporate colonies are not regionated; others display indistinct endozone and exozone regions. Lacking known ovicells, stolons, acanthorods, vibracula, and avicularia, cystoporate colonies may sometimes include mesopores in addition to the ubiquitous cystopores. Many colonies exhibit monticules or maculae or both.

Individual cystoporate zooecia are straight to markedly curved long tubes. Because the zooecia are generally separated from one another by cystopores or vesicles, the walls of nearby zooecia remain distinctly separate and thin throughout the colony. The zooecial walls are calcareous, finely granular in microstructure, and perforated in some species by a few minute pores. The space within each zooecium may be crossed by diaphragms or may lack transverse calcareous partitions. Round or oval in outline, the zooecial aperture lacks an operculum, but often is surrounded by a raised rim (peristome) or partially covered over by a hoodlike lunarium.

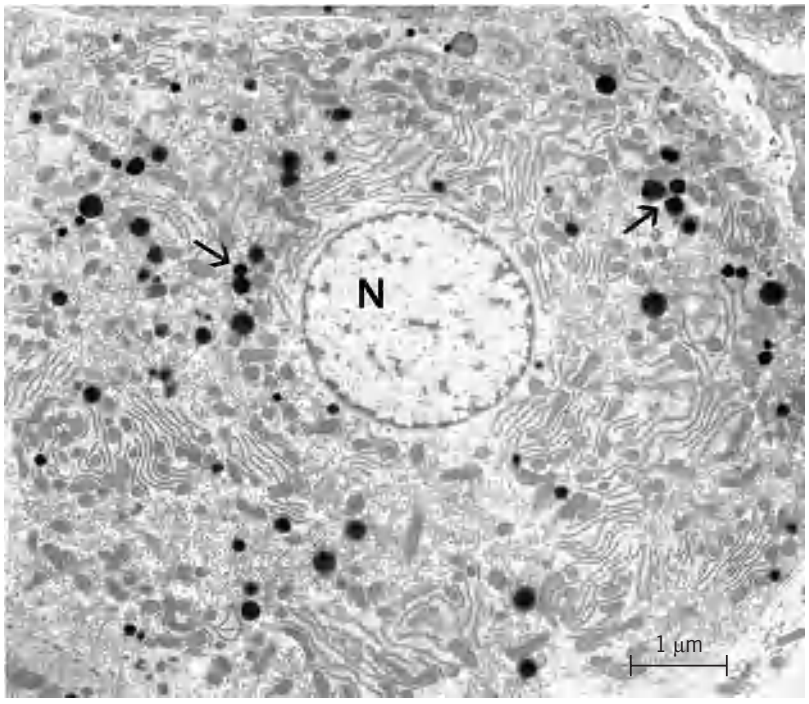
History and classification. The cystoporates may possibly have evolved from the earliest ctenostomes. Cystoporates first appeared late in the Early Ordovician, somewhat earlier than representatives of other stenolaemate orders; thus, the cystoporates may possibly have in turn given rise to those orders, although all these forms may simply share a common ancestor further back in time. Apparently marine, the cystoporates (or ceramoporoids or fistuliporoids; formerly included in the Cyclostomata) became moderately common by Mid-Ordovician time—contributing to building small reefs, as well as to level-bottom communities—and remained so until they died out in the latest Permian. *See* CTENOSTOMATA; CYCLOSTOMATA (BRYOZOA).

Roger J. Cuffey

Cytochemistry

The science concerned with the chemistry of cells. Specifically, the macromolecules of subcellular structures can be treated chemically to form microsopically visible end products. For example, proteins, enzymes, lipids, carbohydrates, and nucleic acids can be directly visualized in cell nuclei, membranes, and organelles by cytochemical methods which generate images that can be viewed by either bright field or light, fluorescence, confocal, or electron microscopes. *See* ELECTRON MICROSCOPE; FLUORESCENCE MICROSCOPE; INTERFERENCE MICROSCOPE; PHASE-CONTRAST MICROSCOPE; SCANNING ELECTRON MICROSCOPE.

Enzyme cytochemistry. Enzyme cytochemical methods detect enzymes associated with subcellular structures (see **illus.**). The method employs a substrate specifically cleaved by an enzyme in the cell to liberate a product which is transformed into a visible precipitate. For example, to localize the enzyme acid phosphatase in lysosomes, a subcellular organelle, a lead sulfide precipitate is produced by splitting a phosphate (product) from cytidine monophosphate (substrate) and transforming the phosphate product initially into lead phosphate and subsequently into lead sulfide; the latter is visible as a brown-black precipitate in the light microscope. Other methods employ azo dye or formazan precipitates rather than lead to produce a visible product. The number of enzymes that can be detected are limited by the availability and specificity of substrates. *See* ENZYME.



Electron micrograph of a hepatocyte from rat liver showing the localization of catalase in peroxisomes by using an enzyme cytochemical method. The catalase activity in peroxisomes was revealed by incubation in a diaminobenzidine medium after fixation with aldehydes. Peroxisomes (arrows) appear as black spherical organelles distributed throughout the cytoplasm. The nucleus is labeled N. (*P. M. Novikoff and A. Yam*)

Immunocytochemistry. Immunocytochemical methods are used to detect specific proteins associated with subcellular structures. The methods are based on the immunologic principle of the built-in specificity of antibody-antigen interactions; that is, an antibody (a protein called immunoglobulin) has a unique binding site that recognizes an antigen. An antigen in a tissue or a cell is detected by employing either a direct or indirect method. The direct method requires a specific antibody linked to a molecule that produces a visible signal; the antibody-molecule complex is applied to a tissue or cells for antigen binding. The indirect method employs an unlabeled primary antibody which is applied to a tissue or cells. The antibody binds to the antigen. The bound primary antibody is then detected by application of a secondary antibody that yields a visible signal. For example, to localize acid mannosidase in lysosomes by the indirect method, a specific antibody (monoclonal or polyclonal) against acid mannosidase (antigen) is applied to a tissue section or to cells and interacts with the intracellular antigen (acid mannosidase) to form a specific antibody-antigen complex, which is then visualized by employing a labeled second antibody that binds to the first antibody. Second antibodies can be labeled with fluorescent probes (such as fluorescein, rhodamine, Texas Red, Cy 3, and Cy 5), enzymes (including peroxidase and alkaline phosphatase), metals (such as gold and ferritin), and high-affinity complexes (for example, avidin-biotin) that are detectable with the appropriate microscope (such as bright-field light, fluorescence, confocal, and electron), in some cases, after application of other reagents to generate a visible product.

Double or triple label immunocytochemistry methods allow multiple antigens to be localized within a tissue or cells. Multiple antigens can be detected at the same time by using different antibodies labeled with probes that fluoresce at different wavelengths; for example, in triple label fluorescence immunocytochemistry one antibody is labeled with fluorescein (520 nanometers wavelength), a second with Cy 3 (570 nm), and another with Cy 5 (670 nm). Multiple antibodies can also be labeled with different enzymes (for example, peroxidase and alkaline phosphatase) or with different size gold particles (such as 5 or 10 nm). When the indirect method is employed, the different primary antibodies used must be from different species so that they can be distinguished using secondary antibodies. Fluorescent probes are detected with conventional fluorescence microscopy or with confocal microscopy using either appropriate filters or lasers. Enzyme probes and gold probes are detected with light and electron microscopy. *See* IMMUNOCHEMISTRY; IMMUNOFLUORESCENCE; PROTEIN.

In-situ hybridization. In-situ hybridization cytochemical methods detect specific deoxyribonucleic acid (DNA) or ribonucleic acid (RNA) sequences in specific cells or intact nuclei, chromosomes, or cytoplasmic subcellular organelles. The method involves the hybridization of known DNA or RNA sequences

(probe) with complementary sequences of DNA or RNA (target) in intact cells. When the target is messenger RNA, the direct product of gene expression, which encodes a specific protein, the purpose of the method is usually to identify the tissues and cells in which a particular gene is expressed. When the target is DNA, the purpose is frequently to map, or localize, a specific gene on a particular chromosome or to detect the presence of a viral genome in infected cells. The binding or hybridization of probe and target is based on the matching of nucleic acid base pairs adenine (A) to thymine (T), and guanine (G) to cytosine (C), within contiguous stretches of nucleic acid sequences. The known probe can be labeled with fluorescent molecules, enzymes, avidin-biotin, or radioactive elements with detection of the probe-target complex accomplished using fluorescence microscopy, cytochemistry, immunocytochemistry, or autoradiography, respectively. Polymerase chain reaction (PCR) also can be performed on the tissue and cells to amplify specific sequences prior to in-situ hybridization to detect single copies of DNA or low levels of RNA. *See* NUCLEIC ACID.

Lectin cytochemistry. Lectin cytochemical methods are employed to detect specific sugar molecules on the plasma membrane and intracellular membranes of cells. Lectins are proteins with binding sites that recognize and attach tightly to a specific sequence of sugar moieties. The lectins exhibit a high degree of specificity for a wide spectrum of sugars. For example, the lectin concanavalin A will detect mannose-rich oligosaccharides such as those in glycoproteins that reside in the endoplasmic reticulum of the cell; wheat germ agglutinin will detect the sugar glucosamine, characteristic of many proteins located in the distal regions of the Golgi apparatus; and peanut agglutinin will detect the sugar galactosamine on cell surfaces. Lectins which are generally isolated from plants do not have enzymatic activity but can be labeled with a variety of molecules (such as enzymes, fluorescent molecules, or metals) allowing their detection by microscopy. *See* CARBOHYDRATE.

Autoradiography. Autoradiographic methods employ radioactive labeled molecules to identify sites within cells where synthesis of macromolecules occurs. This method is based on the property that an isotope emits ionizing particles during radioactive decay. Radioactive labeled precursor molecules (building blocks of molecules) become incorporated into specific cellular sites after injection into living organisms or cells, producing a radioactive product. To establish the location of the radioactive product in tissues or cells, a photographic emulsion is placed over tissues or cells in the dark for a period of time, followed by a photographic developer solution to reveal sites containing the radioactive product. The radioactive product is produced by the interaction of silver bromide crystals suspended in the emulsion, with the ionizing radiation given off by the decaying isotope. The product is made visible by a photographic development solution and appears as black dots or grains. The black dots or grains indicate sites of the radioactive product; these sites are visible in the light and electron microscopes. Autoradiography has determined that DNA synthesis occurs in the nucleus by using a radioactive labeled precursor molecule of DNA tritiated (^{3}H)-thymidine; that collagen is synthesized by fibroblasts by using ^{3}H-proline, an amino acid that is very abundant in collagen; and that thyroglobulin, an iodinated protein from which thyroid hormone (which retains the iodine atoms) is derived by proteolytic degradation, is synthesized by the thyroid and stored in the lumen of thyroid by injecting animals with radioactive iodine. *See* AUTORADIOGRAPHY.

Fluorescence analog. Fluorescence analog cytochemical methods are employed to determine the function of proteins in living cells (for example, during growth, cell division, or migration). A purified protein is coupled to a fluorescent probe and microinjected into living cells, and its fate can be followed in a conventional fluorescent microscope or confocal microscope. For example, fluorescent labeled tubulin, a subunit of microtubules, can be injected into a living cell, and the incorporation of the labeled tubulin into microtubules during cell migration can be followed second by second. Multiple proteins can also be followed in cells by labeling the proteins with a fluorescent probes that emit light of different wavelengths (see **colorplate**).

Measuring pH. Measuring the pH of different organelles involves determining the concentration of hydrogen ions in intracellular organelles in living cells. One method uses fluorochrome probes that fluoresce at different wavelengths depending on the concentration of hydrogen ions. For example, the pH of lysosomes was determined to be 5.0 by using the fluorescent probe fluorescein. *See* PH.

Marker enzymes. Marker enzymes are employed to identify subcellular organelles and to distinguish subcellular organelles from each other in intact tissues and cells and in cell fractions of organelles using enzyme cytochemical methods. For example, acid phosphatase is used as a marker enzyme for lysosomes, catalase for peroxisomes (see illus.), thiamine pyrophosphatase for the Golgi apparatus, nucleoside diphosphatase for the endoplasmic reticulum, and 5′-nucleotidase for the plasma membrane. The marker enzyme concept (that each subcellular organelle possesses one or a set of unique enzymes detectable by enzyme cytochemical methods) has been greatly enhanced by the introduction of immunocytochemistry, which permits the visualization of a larger number of proteins specific for subcellular organelles.

Epitope-tagged proteins. The detection of epitope-tagged proteins introduced by transfection is a molecular genetic-immunocytochemical method to localize proteins for which antibodies are not available. The method involves the construction of an expression vector in which the gene for the protein of interest (for which there is no available antibody) and the DNA sequence encoding an epitope tag (such as a small peptide c-myc protein for which there is an antibody) are linked so that the tandem gene sequences are translated to produce a single protein.

Transfection of the vector into living cells by one of a variety of methods results in production of a fused protein (protein of interest containing the linked epitope tag) in the cells. The fused protein is detected in cells by immunocytochemistry using an antibody to the epitope tag protein (for example, c-myc protein). Commonly used epitope tags include a short peptide-segment from c-myc, a cellular protein normally located in the nucleus, or a peptide from the envelope proteins of the influenza virus. Transfection of living cells is commonly accomplished by an electroporation method which makes the cell membrane transiently permeable to the vector by the application of a brief but intense electric shock.

Histologic dyes. Histologic dye methods use organic dyes to stain cell components in tissue sections of normal and pathologic tissues. The dye methods are used extensively in clinical pathology laboratories for diagnosis because they can be performed easily and quickly. The dyes hematoxylin and eosin are used in a routine staining procedure in pathology laboratories to establish overall architecture or histology of tissues, and they are extremely valuable for determining whether a tissue is normal or pathologic during surgery. Dye methods are available for staining neutral lipids (oil red O or sudan black), for staining nuclear DNA (methyl-green), for staining the nucleolus and ribosomes (pryonin), and for staining acid mucopolysaccharides (alcian blue or ruthenium red). *See* LIPID.

Laser scanning confocal microscopy. Laser scanning confocal microscopy uses lasers to obtain thin optical sections of the entire depth of fluorescent-labeled cells or tissues with an optical system that integrates the focus of the optical section and the focus of illumination to achieve confocality (that is, same focus). This results in the production of fluorescent images of superior clarity because the out-of-focus fluorescence above or below the focused area of a cell or tissue is eliminated. Confocal microscopy is extremely useful for performing double or triple immunofluorescence in which as many as four antigens can be simultaneously viewed: three antigens by fluorescent probes and one by a reflectance probe (such as gold). Images are subsequently processed by computer programs to produce three-dimensional reconstruction, volume rendition, and stereo projections. Study of fluorescent-labeled living cells can be performed optimally by using real-time scanning confocal microscopy, which combines optical sectioning capability of a confocal microscope with rapid scanning and collection of images in real time.

Specimen preparation. Cytochemical methods can be applied to various types of specimens, including cells in culture, sections of tissues, and cell fractions of isolated organelles. For most cytochemical procedures, except those noted above for living cells, the specimen requires fixation with formaldehyde or glutaraldehyde, which functions to keep macromolecules of interest in place. When a tissue is to be used for cytochemical procedures, the tissue is fixed by immersion or perfusion, and then sections are prepared with a cutting instrument (microtome). Depending on the enzyme or antigen to be localized and the level of microscopy (light or electron) used for analysis, tissues can be sectioned without embedding (frozen or nonfrozen) or with embedding (in paraffin for light microscopic pathologic studies or in epon for electron microscopic study). The objective is to preserve the enzyme or antigen of interest as well as the morphology of the tissue or cells. In cases where antigen or enzymes are inhibited by fixation, unfixed frozen sections have to be prepared. The purpose of the experiment will dictate the kind of fixation and sectioning instrument. *See* CELL (BIOLOGY); CELL, SPECTRAL ANALYSIS OF; MICROTECHNIQUE.

Phyllis M. Novikoff

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Cytochrome

A member of a class of proteins containing hemes as prosthetic groups that catalyze oxidation/reduction reactions. Cytochromes are distinct from other heme-containing proteins, such as myoglobin and hemoglobin, which function in the binding and/or transport of molecular oxygen; catalase, which eliminates hydrogen peroxide from cells; and peroxidases, which use hydrogen peroxide to oxidize various substrates.

Types. Cytochromes are identified by the type of heme prosthetic group they contain. Heme a, heme b, and heme c differ from one another by the chemical groups attached to the general iron–protoporphyrin structure (**Fig. 1**). Individual cytochromes can also be further identified by the wavelength of light that the heme group maximally absorbs, which is affected by the protein environment of the heme group. For example, cytochrome b_{562} contains heme b that maximally absorbs light with a wavelength of 562 nm.

Oxidation and reduction. Cytochromes are able to catalyze oxidation/reduction reactions because their heme prosthetic groups can exist in two forms: the oxidized form, in which the iron is in the Fe(III) oxidation state, and the reduced form, in which the iron is in the Fe(II) oxidation state. By switching back and forth between the oxidized and reduced forms, cytochromes can transfer electrons from one substrate to another using the following steps. First, the oxidized form of the cytochrome binds to the reduced

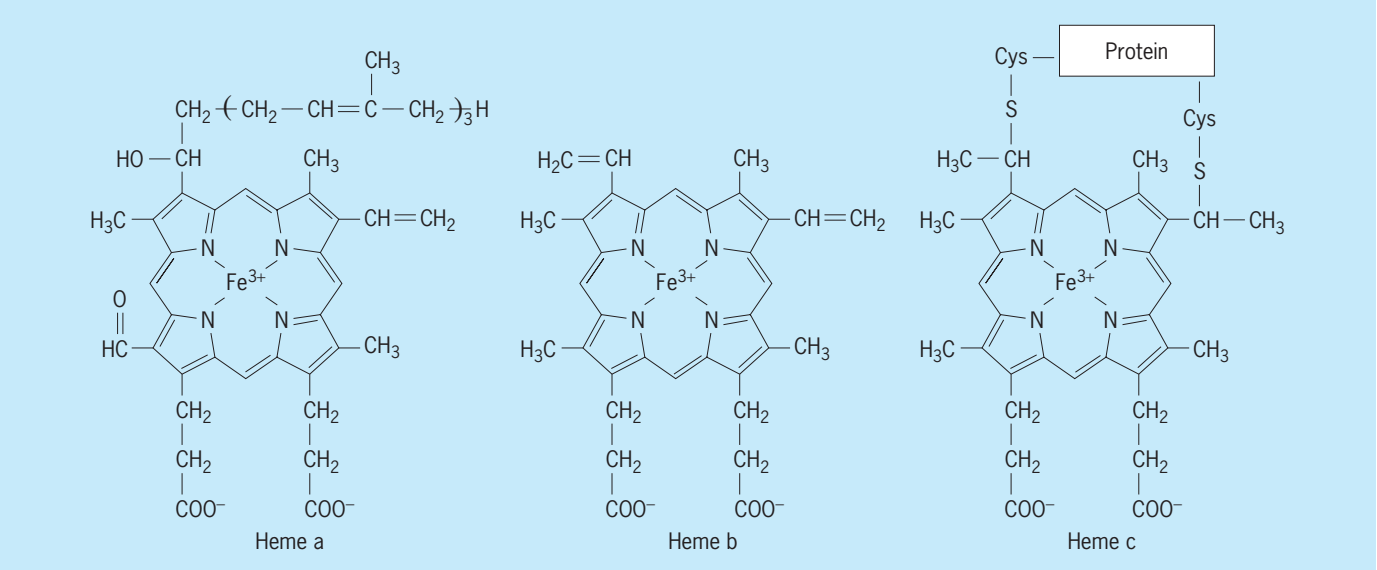

Fig. 1. Structures of the heme a, heme b, and heme c iron-protoporphyrin groups found in cytochromes.

form of one substrate. One electron is then transferred from this substrate to the heme group. This generates the oxidized form of this substrate and the reduced form of the cytochrome. Next, the reduced form of the cytochrome binds to the oxidized form of a second substrate. One electron is then transferred form the heme group to the second substrate. This generates the reduced form of the second substrate and the oxidized form of the cyctochrome.

Electron transport system. Cytochromes play essential roles in oxidative phosphorylation in all organisms except for a few anaerobic bacteria. Oxidative phosphorylation is a very complicated process by which energy released by oxidation/reduction reactions is used to fuel the synthesis of the energy-rich compound adenosine triphosphate (ATP). In the first stage of oxidative phosphorylation, the electron transport system takes the energy released by oxidation/reduction reactions and stores it in the form of a proton gradient across the inner membrane of the mitochondria. In the second stage of oxidative phosphorylation, an enzyme called the ATP synthase takes the energy stored in the proton gradient and uses it to create ATP. Cytochromes are involved only in the first stage of oxidative phosphorylation, the electron transport system that establishes the proton gradient. *See* ADENOSINE TRIPHOSPHATE (ATP).

The electron transport system comprises four protein complexes found within the inner mitochondrial membrane (**Fig. 2**).

Complex I. Complex I is the NADH–coenzyme Q reductase complex. This complex takes electrons from hydrogenated (reduced) nicotinamide adenine dinucleotide (NADH) and transfers them to a membrane-soluble molecule called coenzyme Q. In this process, four protons are transferred across the inner mitochondrial membrane for every two electrons transferred to coenzyme Q. This complex contains no cytochromes. *See* NICOTINAMIDE ADENINE DINUCLEOTIDE (NAD).

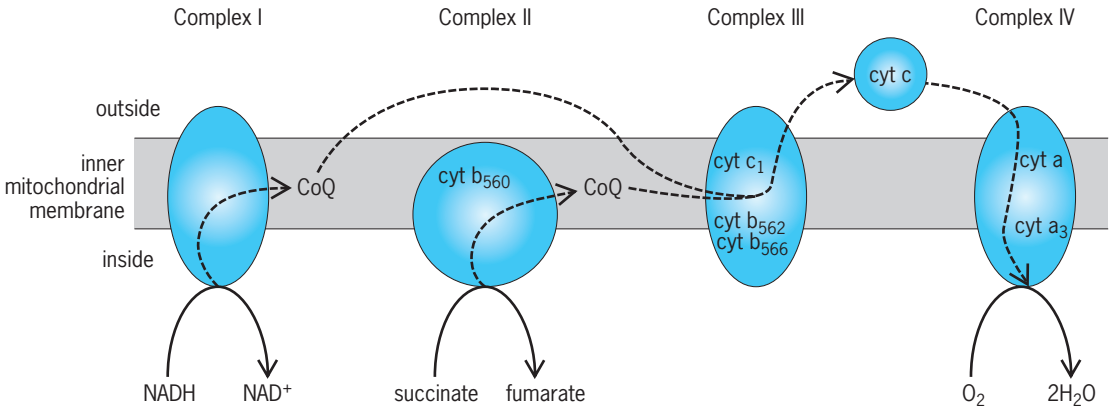

Fig. 2. Diagram of the mitochondrial electron transport system. The dashed arrows represent the flow of electrons as they enter the system either at complex I (from NADH) or at complex II (from succinate) and exit the system at complex IV (to molecular oxygen). The purpose of this process is to pump protons across the inner mitochondrial membrane, establishing a gradient that is used to generate ATP.

Complex II. Complex II is the succinate–coenzyme Q reductase complex. This complex comprises at least four different protein subunits. Part of this complex is the enzyme succinate dehydrogenase, which faces the inner side of the membrane. This enzyme catalyzes the conversion of succinate (a metabolic intermediate found inside the mitochondria) to fumarate, and thereby it transfers electrons from the succinate to an FAD covalently bound to the complex. This complex also contains one 4Fe-4S iron–sulfur cluster, two 2Fe-2S iron–sulfur clusters, and a cytochrome b_{560}. While the path taken by the electrons through this complex is not clear, the electrons ultimately end up transferred to coenzyme Q.

Complex III. Complex III is the coenzyme Q–cytochrome c reductase complex. This complex comprises at least 11 different protein subunits. One protein subunit contains two hemes, b_{562} and b_{566}, another protein subunit contains a 2Fe-2S iron–sulfur cluster, and yet another protein subunit is cytochrome c_1. This complex takes electrons from coenzyme Q. These electrons follow a complex, defined pathway through the iron–sulfur cluster and three heme groups, and ultimately end up transferred to another protein called cytochrome c. In this process, four protons are transported across the inner mitochondrial membrane from inside the mitochondria for every pair of electrons transferred from coenzyme Q to cytochrome c.

Cytochrome c. Cytochrome c is a small cytochrome protein that is not found within the inner mitochondrial membrane. Instead, cytochrome c floats around between the inner and outer mitochondrial membranes. It picks up electrons from complex III and transfers them to complex IV.

Complex IV. Complex IV is the cytochrome c oxidase complex. This complex comprises 13 different protein subunits. One subunit contains a 2Cu-2S copper–sulfur cluster, and another subunit contains two hemes, a and a_3. This complex takes electrons from cytochrome c and ultimately transfers them to molecular oxygen to form water. The electrons travel from cytochrome c to the 2Cu-2S cluster, then to the heme a group, then to the heme a_3 group, and then finally to the bound oxygen molecule. In this process, two protons are transported across the inner mitochondrial membrane from inside the mitochondria for every pair of electrons transferred from cytochrome c to molecular oxygen. *See* BIOLOGICAL OXIDATION; COENZYME; MITOCHONDRIA.

Photosynthesis. Cytochromes also play an essential role in photosynthesis. In bacteria, photosynthesis occurs on the cell membrane at protein complexes called the reaction center, which contains chlorophyll. Light causes electrons from the reaction center to be transferred from the chlorophyll to a quinone molecule. Quinone moves through the membrane and transfers electrons to a cytochrome bc_1 complex. The cytochrome bc_1 complex then transfers these electrons to either a cytochrome c_2 protein (in the case of purple bacteria) or a cytochrome c_{553} protein (in the case of green sulfur bacteria). These soluble cytochrome proteins then return the electrons to the chlorophyll of the reaction center for another cycle. During this process, a proton gradient is established across the cell membrane that is used to fuel ATP generation.

In plants, photosynthesis occurs in the thylakoid membrane of the chloroplasts. In a proton complex called photosystem II, light causes electrons (derived from the oxidation of water to molecular oxygen) to be transferred from chlorophyll to plastoquinones, which move through the membrane and transfer the electrons to the cytochrome b_6f complex. The cytochrome b_6f complex transfers these electrons to plastocyanin, which in turn transfers them to chlorophyll in another protein complex called photosystem I. In photosystem I, light causes these electrons to be passed through another series of carriers until they ultimately reduce $NADP^+$ to NADPH. The NADPH is needed by plant cells to reduce compounds during many important biosynthetic pathways. *See* PHOTOSYNTHESIS.

P-450s. Much recent research has focused on a class of cytochromes called P-450s, which contain the heme b prosthetic group. Their name is derived from the fact that the reduced forms of these proteins (in which the iron in the heme is bound to carbon monoxide) maximally absorb light with a wavelength of 450 nm. In higher animals, all cytochrome P-450s are found within membranes—usually within the membrane of the endoplasmic reticulum but also within the mitochondrial membrane. Cytochrome P-450s are components of monoxygenase systems that catalyze the addition of a hydroxyl group to an organic compound using molecular oxygen. This reaction requires electrons passed to cytochrome P-450 from NADH by a protein called cytochrome P-450 reductase.

The human genome encodes 57 cytochrome P-450 proteins, each thought to be specific for hydroxylating a different organic compound. Many of these cytochrome P-450s are involved in hydroxylating xenobiotics (compound such as drugs or toxins that are not ordinarily found within the cell). For example, the carcinogen benzo[*a*]pyrene found in cigarette smoke is hydroxylated and detoxified by a cytochrome P-450. Other cytochrome P-450s are involved in the synthesis of bile acids and steroid hormones. Some are involved in fatty acid metabolism, particularly the omega oxidation of fatty acids, in which the methyl carbon at the end of the fatty acid hydrocarbon chain is hydroxylated. The biological functions of many cytochrome P-450s encoded in the human genome have still not been determined.

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Cytokine

Any of a class of soluble proteins (molecular weight less than 100,000 daltons, or 100 kD) that are released by a cell to send messages that are

delivered to the same cell (autocrine), an adjacent cell (paracrine), or a distant cell (endocrine). The cytokine binds to a specific receptor and causes a change in function or in development (differentiation) of the target cell. Cytokines are involved in reproduction, growth and development, normal homeostatic regulation, response to injury and repair, blood clotting, and host resistance (immunity and tolerance). Unlike cells of the endocrine system, many different types of cells can produce the same cytokine, and a single cytokine may act on a wide variety of target cells. Further, several cytokines may produce the same effect on a target, so the loss of one type of cytokine may have few if any consequences for the organism; this situation is called redundancy. Finally, the response of a target cell may be altered by the context in which it receives a cytokine signal. The context includes other cytokines in the milieu and extracellular matrix. Thus has developed the concept of cytokines as letters of an alphabet that combine to spell words which make up a molecular language.

Cytokines may be divided into six groups: interleukins, colony-stimulating factors, interferons, tumor necrosis factor, growth factors, and chemokines.

Interleukins. Interleukins are proteins that are produced primarily by one type of lymphocyte (a type of white blood cell, or leukocyte) or macrophage (a large phagocytic cell of the immune system) and that act on other leukocytes. However, it is now known that production of interleukins is not confined to lymphocytes and macrophages. Interleukins may be grouped into families with similar structure and function. The plethora of interleukins suggests considerable redundancy, although it is suspected that each interleukin may have some special purpose. Thirty-four types are described below.

Interleukin (IL)-1α is composed of two protein subunits produced by macrophages, a specialized antigen-presenting cell called a dendritic cell, B (bone marrow–derived) and T (thymus-derived) lymphocytes (also known as B and T cells). IL-1β is also a dimer of two proteins and is produced by B and T cells, macrophages, dendritic cells, skin epithelium, fibroblasts, blood vessel endothelium, joint-lining cells, cartilage cells, bone marrow and circulating leukocyte cells, liver cells, adrenal gland cells, pancreatic islet β cells (which secrete insulin), fertilized eggs (early embryos), and cells of the nervous system. IL-1α and IL-1β bind to the same receptor, which may occur on a variety of target cells, for example, T lymphocytes (stimulating immunity), those of the nervous system (causing fever), macrophages and dendritic cells, and endothelial cells (activating inflammation and blood clotting). Soluble IL-1 receptors have been identified and synthesized. They antagonize IL-1 action, and can prevent shock in response to bacterial endotoxin. *See* CELLULAR IMMUNOLOGY.

Interleukin-2 is a glycoprotein produced by a subset of T lymphocytes called T-helper type 1 cells (Th1), which "help" in immune responses. IL-2 is a prototypic example of a cytokine produced by Th1 lymphocytes. Th1-type cytokines promote cell-mediated immunity and inflammation (so IL-1 is included in the Th1 group); they inhibit generation and production of Th2-type cells and Th2-type cytokines. Conversely, Th2 cytokines inhibit Th1 responses. Hence, there is reciprocal inhibition and Th1/Th2 cytokine balance. IL-2 may also be produced by B cells, natural killer cells, and oligodendrocytes (which myelinate nerve cells) in the brain. IL-2 stimulates the development of cytotoxic (killer) T lymphocytes and B lymphocytes (which develop into antibody-secreting cells). IL-2 also activates nonspecific (unlike T and B cells) cytotoxicity in natural killer cells. Natural killer cells participate in the early defense response to infection and may reject certain types of cancers early in their development. IL-2 may also be made by fetal trophoblast cells that lie at the feto-maternal interface in the placenta. Receptors for IL-2 are found on lymphoid cells and on astroglia of the central nervous system.

Interleukin-3 is a growth factor for multiple lineages of hemopoietic cells (which develop into blood cells). It was originally defined by its ability to stimulate the growth of mast cells (a large leukocyte that stores various inflammatory mediators). IL-3 is produced by Th2 cells, as well as by nonlymphoid cells (including natural killer cells, mast cells, endothelial cells, and skin keratinocytes) and can act as an immunoregulatory factor. For example, IL-3 (a Th2 cytokine) prevents IL-2 (a Th1 cytokine) from activating killer activity in natural killer cells.

Interleukin-4 is also a product of Th2 T cells, but may be produced by nonlymphoid cells such as fetal trophoblast. IL-4 promotes antibody responses by B lymphocytes, inhibits cell-mediated immunity and natural killer cell activation by IL-2, and contributes to granuloma formation (possibly by acting on macrophages and endothelial cells). Granulomas are largely composed of macrophages and represent a particular type of chronic inflammation that can lead to tissue scarring and, hence, disease.

Interleukin-5, another Th2-type cytokine, promotes the development of B cells which ultimately produce immunoglobulin E, a class of antibody involved in immune responses to certain parasites and in allergies such as asthma and hay fever. IL-5 may arise from eosinophils and basophils (a type of granulocyte) and also stimulates the bone marrow to produce eosinophils, which are involved in allergies; eosinophils contain a 12-kD major basic protein which is highly toxic to certain parasites.

Interleukin-6, also known as β-2 interferon and liver cell growth-stimulating factor, is a Th2-type cytokine produced by a wide variety of cell types including T cells, B cells, and macrophages. IL-6 contributes to B lymphocyte (antibody) responses and stimulates the liver to produce protease inhibitors (such as α1-antirypsin and acute phase proteins) which dampen inflammation. IL-6 may also stimulate fetal trophoblast cell function.

Interleukin-7 is a polypeptide that acts as a growth and maturation factor for precursors of B and T

lymphocytes. IL-7 is produced by stromal cells in the thymus as well as by bone marrow cells, and is particularly important for the development of a type of T cell at mucosal surfaces which express $\gamma\delta$ T cell receptor (TcR) instead of the usual $\alpha\beta$-type receptor that is found on most circulating T cells and T cells in the lymph nodes and spleen. The $\gamma\delta$ TcR allows the T cell to recognize certain types of antigens at mucosal surfaces which pose a threat. Recognition can occur without the need for antigen presentation by a macrophage or dendritic cell, which present the antigen as a peptide fragment in the groove of conventional class I or class II major histocompatibility complex (MHC)-self molecules. Some antigenic peptides are recognized via $\gamma\delta$ TcR in association with atypical class I MHC-type antigens. IL-7 also enhances the generation of IL-2-activated nonspecific killer cells and IL-2 activated antigen-specific cytotoxic T cells, and thus is viewed as a Th1-type cytokine.

Interleukin-8 is actually a group of peptides produced by a variety of cell types, including vascular endothelial cells, macrophages, and fibroblasts. IL-8 activates and recruits polymorphonuclear leukocytes in the inflammatory process, and is involved in the initiation of labor and delivery in pregnant women. IL-8 is not found in the mouse.

Interleukin-9 is produced by a subset of Th2 cells and stimulates hemopoietic cells that develop into red blood cells. Mast cells and B cells may also be a target.

Interleukin-10 is produced by the Th2 subset of T cells as well as by B lymphocytes, macrophages, skin keratinocytes, nonepithelial nonleukocytic cells (stroma) in the endometrium, and fetal trophoblast cells that form the placenta. IL-10 inhibits secretion (and function) of proinflammatory macrophages and activation of vascular endothelium and polymorphonuclear leukocytes (eosinophils, basophils, and neutrophils). Hence, inflammation and clotting are suppressed. IL-10 protects against lethal shock triggered by bacterial endotoxin in sepsis by inhibiting production of proinflammatory (Th1) cytokines and other factors that contribute to low blood pressure.

Interleukin-11 is a factor produced by bone marrow–derived fibroblasts and supports the growth and development of certain types of cells, such as B lymphocytes, neutrophils, and platelet-producing megakaryocytes.

Interleukin-12 is a Th1-type produced by macrophages and some dendritic cells. The p20 form of IL-12 stimulates Th1 lymphocyte development and activation on natural killer–type cells that produce interferon γ. IL-12 may substitute for IL-2 in Th1-type responses. The p40 form of IL-12 acts on macrophages.

Interleukin-13 is produced by activated Th2 T cells as well as by natural killer cells and mast cells. IL-13 inhibits proinflammatory cytokine production induced by bacterial endotoxin. It binds to one of the two types of receptors for IL-4, suppresses production of tumor necrosis factor (TNF)-α, is a chemoattractant for vascular endothelial cells and eosinophils (which it stimulates), and promotes B lymphocyte production of IgE, which is important in allergy.

Interleukin-14 is a Th2-type cytokine produced by microvascular endothelial cells (along with IL-3,7,8,11,15, tumor necrosis factor-α and transforming growth factor-β). It may also be produced by T and B cells, and like other Th2-type cytokines, stimulates B cells.

Interleukin-15 is a Th1-type IL-2-like cytokine that binds to IL-2 receptor but is made primarily by monocytes/macrophages, fibroblasts in joint lining (synovia), and keratinocytes in inflammatory responses. Brain astrocytes and microglia stimulated by IL-1, interferon (INF)-γ and TNF-α may produce IL-15. It is also stimulated by IL-10, and activates INF-γ and IL-4 production in lymphoid cells within mucosal epithelia. It can activate natural killer cells, T lymphocytes, B lymphocytes, and neutrophils (that then produce IL-8).

Interleukin-16 is a proinflammatory (Th1) cytokine produced by T cells, fibroblasts, eosinophils, and mast cells. It stimulates migration of $CD4^+$ T cells (that recognize antigen in association with Class II MHC), upregulates (increases surface expression of) receptors for IL-2, stimulates monocytes, eosinophils, and precursors of B cells.

Interleukin-17 is a Th1-type cytokine produced almost exclusively by memory T cells (defined by their $45RO^+$ surface marker). IL-17 induces production of proinflammatory Th1-type cytokines by a variety of cells, and is involved in angiogenesis (origin and development of blood vessels) and recruitment of neutrophils to tissues.

Interleukin-18 is a Th1-type cytokine similar to IL-12. It is produced by macrophage-type cells and skin cells, and acts to stimulate production of interferon γ by T lymphocytes and natural killer cells.

Interleukin-19 is a product of B cells and macrophages and can act on T cells and macrophages to increase production of IL-4,5,10,13 and TNF-α.

Interleukin-20 is produced by macrophages and skin keratinocytes and acts on keratinocytes.

Interleukin-21 is derived from T cells and stimulates T and natural killer cells.

Interleukin-22 is a Th1-type cytokine produced by Th1 cells and macrophages that stimulates inflammation.

Interleukin-23 is a product of dendritic cells that acts on T cells.

Interleukin-24 is a Th2 type cytokine that can also derive from macrophages, natural killer cells, and fibroblasts. IL-24 stimulates megakaryocytes.

Interleukin-25 is a Th2 cytokine produced from nonlymphoid stroma. It can stimulate certain bone marrow cells not committed to a particular differentiation pathway.

Interleukin-26 is derived from T and natural killer cells and stimulates epithelial cells to make IL-8 and IL-10.

Interleukin-27 is a product of dendritic cells that acts on natural killer and T cells.

Interleukin-28A and B are interferon-related

molecules produced by dendritic cells and exert an antiviral effect.

Interleukin-29 is similar to IL-18A and IL-18B.

Interleukin-30 is a product of antigen-presenting cells (for example, macrophages and dendritic cells) that acts on natural killer and T cells.

Interleukin-31 is a T cell–derived factor that acts on macrophages and epithelial cells.

Interleukin-32 is a product of T and natural killer cells that stimulates macrophages to make TNF-α, IL-8, and other proinflammatory proteins.

Fibroleukin is a soluble fibrinogen-related protein produced by a subpopulation of T cells and by a wide variety of other cell types. Fibroleukin appears to alter maturation of dendritic cells; therefore, its role may be to regulate (that is, suppress) immune responses. It has not yet been given an interleukin number. A membrane-associated form of fibroleukin can act directly as a prothrombinase, leading to clotting. *See* CELLULAR IMMUNOLOGY; FIBRINOGEN; HISTOCOMPATIBILITY; INFLAMMATION.

Colony-stimulating factors. Colony stimulating factors (CSFs) are produced by lymphoid and nonlymphoid cells. These factors provide a mechanism whereby cells that are distant from bone marrow can call for different types of hemopoietic progeny. There are also growth-promoting actions of locally produced CSFs within the bone marrow to stimulate progenitors to differentiate into macrophages (CSF-M = CSF-1), granulocytes (CSF-G), or colonies containing both cell types (CSF-GM). Generation of erythrocyte colonies is stimulated by erythropoietin (EPO), which is produced by the kidney. Certain interleukins may affect particular steps in the growth and differentiation pathways activated by CSFs. Macrophage-lineage progenitors of specialized antigen-presenting dendritic cells that activate or suppress responses of T and B lymphocytes; proliferation of dendritic cell precursors is driven by the cytokine flt3 ligand. CSF-1 may promote Th2 over Th1 type responses. CSF-G is administered to patients receiving cytotoxic anticancer drugs to enhance recovery of blood neutrophils that fight infection. In the uterus, CSFs are produced by epithelium and stromal cells, with epithelium dominating. CSF-1 enhances trophoblast growth, probably via stimulation of macrophages. CSF-GM appears to have a direct effect on implantation, and stimulates preimplantation division of embryonic cells within its shell of zona pellucida. *See* EMBRYONIC DIFFERENTIATION; FERTILIZATION.

Interferons. Interferons classically interfere with the virus-replication mechanism within cells. Interferon-α (produced by leukocytes) and interferon-β (produced by fibroblasts) activate cytotoxicity in natural killer cells. Interferon γ also activates natural killer cells and is a potent activator of macrophages as well. In certain domestic animal species, fetal trophoblast cells of the placenta produce novel types of interferon (interferon tau), which signals the mother to continue to make steroid hormones necessary for the pregnancy to continue. In primates, such as the human, this signaling is mediated by chorionic gonadotropin. IL-28A, and IL-28B and IL-29 have been designated as IFN-λ1, -λ2, and -λ3.

Tumor necrosis factors. Tumor necrosis factor (TNF)-α is produced by a variety of cell types, but activated macrophages represent the dominant source. TNF-α activates natural killer cell cytotoxicity, enhances generation of cytotoxic T lymphocytes (if there is inhibition by transforming growth factor-β, as described below), and activates natural killer cells to produce interferon γ. TNF-α also acts on vascular endothelium to promote inflammation and thrombosis (formation of a blood clot in a blood vessel); these effects may be enhanced by the copresence of IL-1 and interferon γ. TNF-α may also induce apoptosis (programmed cell death), and this also may be enhanced by copresence of interferon γ. Indeed, a synergistic effect of TNF-α and IFN-γ has been described in a variety of disorders; for the cytokines to act (in some cases), a Toll-like receptor (TLR) signal is also required. Toll-like receptors are related to the fruit fly protein *Toll*; they enable cells of the innate immune system to recognize "danger," for example, TLR4 binds bacterial endotoxin.

At low levels, TNF-α may serve as a growth and differentiation signal to trophoblasts and other cell types. Soluble receptors to TNF-α may bind and neutralize biological function. TNF-β is a product of Th1 T cells; in addition to providing help in proinflammatory cell–mediated immune responses, these cells produce delayed-type hypersensitivity reactions, in which macrophages are locally recruited and activated to kill intracellular pathogens, for example, certain bacteria. TNF-β has interferon-type activity and a narrower spectrum of action than TNF-α. Originally, TNF-β was called lymphotoxin as it was produced by lymphocytes and could attack certain types of cancer cells.

Other growth factors. Transforming growth factors (TGFs) have the ability to promote unrestrained proliferation of cells that are otherwise benign. TGFs have therefore been implicated in development of cancer. There are two groups of TGFs, TGF-α and TGF-β. TGF-α is a 5-kD peptide produced by a variety of cells. It collaborates with TGF-β, a 25-kD peptide, in promoting unrestrained tumorlike growth. Fibroblasts with a benign phenotype must attach to a surface to grow in culture (usually to a plastic disk), but in the presence of TGF-α and TGF-β, some of the cells (if suspended in an agar gel) will grow into colonies. This "anchorage-independent" growth is an identifying property of transformed/malignant cancer cells.

Epidermal growth factor, produced by the salivary gland, can substitute for TGF-α. TGF-β has potent pleiotropic effects on a wide variety of tissues and is a potent fibrogenic and immunosuppressive agent. Classical TGF-βs in mammals are divided into β1, β2, and β3, and have some functional differences. For example, TGF-β2 is less inhibitory to developing bone marrow cells compared with TGF-β1, and TGF-β3 may inhibit scar formation (fibrogenesis) induced by TGF-β1. There are other TGF-βs produced by tumor

cells and found in the uterus in pregnancy that differ in molecular size and certain functional properties.

More distantly related molecules such as activin and inhibin are involved in regulation of the endocrine and reproductive systems. Insulin, produced by β cells in the islets of the pancreas, is a growth factor for many cells, and belongs to the endocrine system. Related are insulinlike growth factors (somatomedins) which may be produced by a variety of cell types. A distinct fibroblast growth factor and platelet-derived growth factor have also been defined. *See* GROWTH FACTOR; TUMOR.

Chemokines. Chemokines are chemoattractant cytokines of small size (7–14 kD); they are subdivided into four families: CXC, CC, C, and CX_3C. Chemokines are produced by macrophages stimulated by bacterial endotoxins, and they control the nature and magnitude of cell infiltration in inflammation. They may affect T cell subsets, monocytes, natural killer cells, basophils, neutrophils, eosinophils, dendritic cells, and mast cells. The mouse chemokine known as MIP-1α is the functional equivalent of IL-8 in the human.

Wound healing. Wound healing is probably the most common phenomenon in which the importance of cytokines is seen. Cytokines ensure that the restorative sequences are carried out in the appropriate order by signaling blood cells and vascular endothelium to coagulate and fill in a wound opening, recruiting and signaling macrophages and neutrophils to engulf microbes, and guiding protective skin epidermal cells to grow over the wounded area (a TGF-β effect). If the damage is more extensive, cytokines stimulate production of new skin cells, blood vessels, connective tissue, and bone. It has been suggested that TGF-β, which inhibits certain enzymes (proteases) that break down tissues, may promote more rapid healing of surgical wounds. Such treatment could reduce the postsurgery recovery period of many individuals and animals. Inhibiting excessive scar formation will be beneficial in plastic surgery. Manipulating the levels of specific cytokines in the body is likely to enhance our ability to treat many different types of illnesses in the future. The immunosuppressive cytokines may be particularly helpful in more successful transplantation of organs (kidney, heart, liver, skin). *See* IMMUNOSUPPRESSION.

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Cytokinins

A group of plant hormones (phytohormones) that, together with other plant hormones, induces plant growth and development. The presence of an unknown compound in the phloem that stimulated cell proliferation of potato tubers was demonstrated in the early 1900s by G. Haberlandt. In the 1940s J. van Overbeek found that the milky endosperm from immature coconuts is rich in compounds that promote cell proliferation. Then, in 1954, during investigations of the nutritional requirements of tobacco callus tissue cultures, F. Skoog and C. O. Miller discovered a very active substance that stimulated cell division and was formed by the partial breakdown of aged or autoclaved herring sperm deoxyribonucleic acid (DNA). The compound was named kinetin because it induced cytokinesis of cultured tobacco cells. Kinetin has not been found in plants, and it can be formed from a mixture of autoclaved furfuryl alcohol and adenine. In 1961 Miller obtained an active substance that stimulates cell division from the milky endosperm of corn (*Zea mays*) seeds; in 1964 D. S. Letham, identified this active compound and named it zeatin. Cytokinin is now the accepted term for those naturally occurring and synthetic compounds that promote the cell division (cytokinesis) and growth of various plant tissues.

Analogs. The discovery of kinetin stimulated the synthesis of over 150 different species of kinetin analogs. Tobacco pith tissue culture bioassays indicate that most of the highly active cytokinins are N6-substituted adenines containing 5-C alkyl side chains (**Fig. 1**). Two of the most potent synthetic cytokinins are kinetin and N6-benzyl adenine. *N*,*N*′-diphenylurea and its derivatives are also considered cytokinins. These compounds induce cell division and growth but are yet to be confirmed as naturally occurring cytokinins.

Since the isolation of the first naturally occurring cytokinin, zeatin, more than 25 different cytokinins have been isolated from plants. A few cytokinin bases such as isopentenyladenine and *cis*-zeatin have been found in transfer ribonucleic acid (tRNA) of various organisms. The most active form of cytokinin is *trans*-zeatin isomer; the *cis* isomer of zeatin, on the other hand, shows little cytokinin activity.

Each cytokinin can exist as a free base, that is, a nucleoside in which a ribose group is attached to the nitrogen atom of position 9 or a nucleotide in which phosphate is esterified to the 5′ carbon of ribose. In most cases, bases have been shown to be more active than their corresponding nucleosides or nucleotides in bioassays. However, it is still not clear

what the number of active forms of cytokinins may be, because cytokinin bases can be converted to the corresponding nucleosides and nucleotides in plant cells within a short period of time. *See* BIOASSAY.

A group of cytokinin analogs has been reported to be highly potent inhibitors of the cytokinin-requiring callus tissues. On the basis that growth inhibition was abolished or reduced in the presence of higher concentrations of cytokinin, these analogs are named anticytokinins or cytokinin antagonists. One of the most potent anticytokinins is 3-methyl-7-(pentylamino)pyrazolo[4,3-*d*]-pyrimidine (**Fig. 2**).

Biosynthesis and metabolism. Since certain tRNA species contain cytokinins, it has been suggested that turnover of tRNA serves as a source of free cytokinins, but it is now generally accepted that tRNA turnover is not a major source of free cytokinin under normal conditions. Indeed, it has been demonstrated that plant tissues contain enzymes that are capable of the direct synthesis of free cytokinins. The enzyme Δ^2-isopentenylpyrophosphate: AMP-Δ^2-isopentenyl- transferase, or isopentenyltransferase, which has been isolated from various organisms, catalyzes the formation of Δ^2-isopentenyladenosine-5′-monophosphate, a cytokinin nucleotide, from Δ^2-isopentenylpyrophosphate and adenosine-5′-monophosphate. Once the cytokinin nucleotide is formed, other enzymes in plant cells convert it to the corresponding nucleoside and base. Oversynthesis of cytokinins in plant tissue causes abnormal growth: crown gall tumor disease caused by the bacterium *Agrobacterium tumefaciens* is an example of excessive production of cytokinins in local tumor tissue. Tissue from crown gall tumors can grow on a simple medium lacking plant hormones because the tumor tissue overproduces both cytokinin and auxin. This is due to the insertion of a piece of bacterial plasmid DNA into the plant nuclear genomes causing activation of the gene responsible for the regulation of cytokinin production. Roots have been shown to be the major site of cytokinin biosynthesis, but stems and leaves are also capable of synthesizing cytokinins. It is possible that all actively dividing cells are capable of cytokinin biosynthesis. *See* CROWN GALL.

Cytokinins applied to plant tissues are rapidly converted to a variety of metabolites by enzymes which catalyze side-chain modifications, side-chain cleavage, or conjugation of cytokinins with sugars or amino acids. The formation of glycosides and amino acid conjugates is a very common metabolic response of plant tissues to externally applied cytokinins.

Detection. The qualitative and quantitative aspects of cytokinin extracted from plant tissues vary with the types of extraction solvent and the procedure employed, but, in general, the purification and identification of naturally occurring cytokinins has been facilitated by the development of high-performance liquid chromatography, gas chromatography–mass spectrometry, and monoclonal antibodies. The levels of the major cytokinins in plant tissues range from 0.01 to 150 nanomoles per 100 g of fresh tissues.

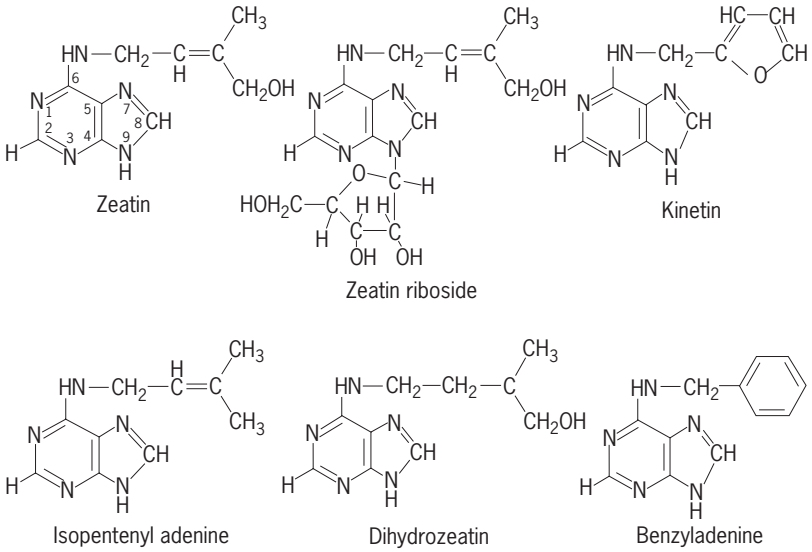

Fig. 1. Structures of naturally occurring (zeatin, zeatin riboside, isopentenyl adenine, and dihydrozeatin) and synthetic (kinetin and benzyladenine) cytokinins. All of these compounds are adenine derivations in which the purine ring is numbered as shown for zeatin. Zeatin can exist in the trans (as shown) or cis (with CH_3 and CH_2OH groups interchanged) configuration.

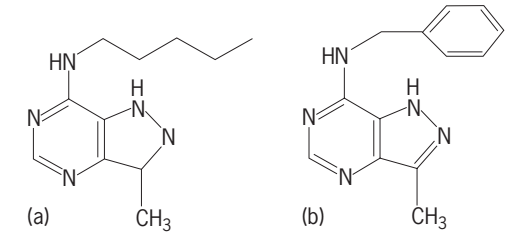

Fig. 2. Structures of two synthetic anticytokinins: (*a*) 3-methyl-7-(pentylamino)pyrazolo(4,3-*d*)pyrimidine and (*b*) 3-methyl-7-(benzylamino)pyrazolo(4,3-*d*)pyrimidine.

However, most of the reported cytokinin levels in plant tissues are rough estimates, because the various forms of cytokinins are not accurately measured. Certain cytokinins are also present in tRNAs of plants, animals, and microorganisms.

Roles of the cytokinins. Cytokinins exhibit a wide range of physiological effects when applied externally to plant tissues, organs, and whole plants. Exogenous applications of this hormone induce cell division in tissue culture in the presence of auxin. The formation of roots or shoots depends on the relative concentrations of auxin and cytokinin added to the culture medium. High auxin and low cytokinin concentrations lead to root formation, while low auxin and high cytokinin concentrations give shoots. Tissue culture techniques have been employed by plant biotechnologists to grow genetically engineered plant cells into whole plants. Cytokinins appear to be necessary for the correlated phenomena of mitosis and nucleic acid synthesis. Cytokinins delay the aging process in detached leaves by slowing the loss of chlorophyll. Retention of chlorophyll is accompanied by chlorophyll synthesis. Exogenously applied cytokinins are effective on fresh flowers, vegetables, and fruits as post-harvest preservatives; however, the use of cytokinins for the preservation of plant materials is not practiced widely.

Cytokinin effects also include breaking of dormancy, promotion of seed germination, stimulation and nutrient mobilization, enhanced anthocyanin and flavanoid synthesis, increased resistance to disease, and stimulation of the opening of stomates. Under some conditions, cytokinin enhances shoot branching and the expansion of cells in leaf disks and cotyledons. In a mammalian system, cytokinin nucleotides have been shown to suppress leukemic cell growth, but the mechanism by which this occurs is unknown. Certain cytokinins located adjacent to the anticodon of tRNA have been shown to influence the attachment of amino acyl tRNA to the ribosome–messenger ribonucleic acid (mRNA) complex and may be involved in protein synthesis. *See* AUXIN; DORMANCY; PLANT PHYSIOLOGY; TISSUE CULTURE.

Cytokinin-binding molecules. It is a widely held view that there are macromolecules such as proteins which interact between cytokinin molecules and the cytokinin-responsive genes. According to this hypothesis, the initial event of hormone action involves the binding of a hormone to a receptor, with the hormone–receptor complex then binding to a specific region of gene. There have been numerous papers reporting the isolation of cytokinin-specific and relatively high-affinity cytokinin-binding proteins. However, the physiological role of any of the reported cytokinin-binding proteins is unknown, and these proteins cannot yet be considered to be cytokinin receptors.

Cytokinin-modulated gene expression. The mechanism of action of cytokinin on plant growth and development is poorly understood. It has been demonstrated that specific proteins are induced, enhanced, reduced, or suppressed by the hormone. Some of the enzymes or proteins regulated by cytokinins have been identified. The activities of acid phosphatase, ribulose 1,5-bisphosphate carboxylase, invertase, light-harvesting chlorophyll *a*/*b*-binding protein, and hydroxypyruvate reductase are all enhanced by the hormone, but the details of how these enzymes are stimulated by cytokinins remain to be studied. Applications of cytokinin to cytokinin-responsive tissues also regulate the synthesis of specific mRNAs; some of the mRNAs are induced and some are suppressed. By using modern techniques, plant molecular biologists are able to show that mRNA changes occur within 60 min of cytokinin application to plant cells. Initial evidence suggests that cytokinins regulate, at least in part, the transcriptional process of gene expression by turning on or off specific genes and stimulating or suppressing the synthesis of specific mRNAs. Another possible action of cytokinins is the regulation of a posttranscriptional process such as stabilization of mRNA. Scientists also demonstrated that cytokinins specifically increase the rate of protein synthesis and the effect seems to be on the translation of mRNA into proteins. *See* PLANT GROWTH; PLANT HORMONES.

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Cytolysis

An important immune function involving the dissolution of certain cells. There are a number of different cytolytic cells within the immune system that are capable of lysing a broad range of cells. The most thoroughly studied of these cells are the cytotoxic lymphocytes, which appear to be derived from different cell lineages and may employ a variety of lytic mechanisms. Cytotoxic cells are believed to be essential for the elimination of oncogenically or virally altered cells, but they can also play a detrimental role by mediating graft rejection or autoimmune disease. There are two issues regarding cytotoxic lymphocytes that are of concern: one is the target structure that is being recognized on the target cell, that is, the cell that is killed, which triggers the response; and the other is the lytic mechanism. *See* CELLULAR IMMUNOLOGY.

Cytotoxic lymphocytes. When freshly isolated, large granular lymphocytes from peripheral blood are tested in cytotoxicity assays, they spontaneously lyse certain tumor cells. These cytotoxic cells are called natural killer cells, and they are important mediators of innate immunity as a first line of defense against invading pathogens. They are unique in that no previous sensitization is required for them to kill. It now appears that a number of different receptors on natural killer cells are capable of activating the lytic machinery. Recently, it has been found that these cells also express inhibitory receptors that actually inhibit cell lysis, thus adding another level of complexity to the regulation of cytolysis by these cells.

Another killer cell, called the lymphokine-activated killer cell, is a product of culturing large granular lymphocytes in relatively high concentrations of certain growth factors. They lyse any target cell, including cells from freshly isolated tumors, and are being employed in cancer therapy. Lymphokine-activated killer cells may also be important in mounting a vigorous response under conditions of extreme immunological stress. Very little is known about the mechanisms by which these cells recognize and lyse the target cell.

The last group of cytotoxic cells is the cytotoxic T lymphocyte. These are T cells that can lyse any target cell in an antigen-specific fashion. That is, as a population they are capable of lysing a wide range of target cells, but an individual cytotoxic T lymphocyte is capable of lysing only those target cells which bear the appropriate antigen. Cytotoxic T lymphocytes recognize peptide antigens in the context of major

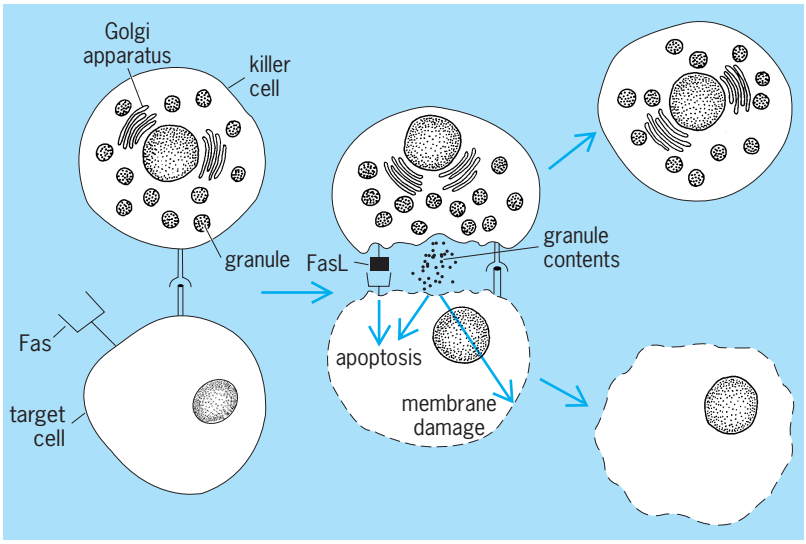
histocompatibility complex class I antigens. These are truly immune cells in that they require prior sensitization in order to function. These cells are thought to mediate graft rejection, mount responses against viral infections and intracellular bacterial infections, and play a major role in tumor destruction.

Cytolytic cycle. Cytotoxic cell-mediated lysis is divided into three distinct steps (see **illus.**). The first step is conjugation, when the killer cell determines if the target cell expresses the appropriate antigen and binds to it via a complex array of adhesion molecules. The second step involves the programming for lysis in which the lytic event is triggered. The third step is the destruction of the target cell. This step occurs after the cytotoxic T lymphocyte has broken free from the target cell. The cytotoxic cell is then free to repeat multiple cycles of killing. The lytic cycle is extremely rapid, and the process can be completed within 20–30 min after mixing the target with the killer cell.

Any proposed explanations of the mechanism of cytolysis must take the above steps and the following observations into consideration. It is clear that direct cell contact between the target cell and the killer cell is absolutely required for initiating the lytic mechanism; if cell contact is prevented, no killing occurs. Another interesting aspect of this process is that the killer cells remain unscathed during the lytic event, suggesting that the killer cell must either employ a unidirectional lytic mechanism or be resistant to the lytic mechanism. Also, when many, but not all, target cells die after cytotoxic T-lymphocyte interaction, nuclear damage with rapid deoxyribonucleic acid (DNA) fragmentation precedes detectable plasma membrane damage. *See* ANTIGEN; HISTOCOMPATIBILITY.

It is becoming clear that cytotoxic lymphocytes employ multiple mechanisms designed to initiate target cell destruction. If cytolysis exists to protect the organism from invading pathogens, there should be redundancies in the system so that, if the pathogen has a mechanism for escaping one cytolytic pathway, alternative mechanisms would still be functional. For example, some viruses have evolved strategies to evade the immune system, including the specific inhibition of certain cytolytic pathways. Therefore, there may be a number of ways in which killer cells can trigger cell death in the target cells that have yet to be recognized. Two mechanisms, which are the best studied to date, are the degranulation of cytolytic granules and the triggering of death receptors found on target cells.

Cytolytic granules. In the degranulation mechanism of killing, cytotoxic cells release the contents of cytotoxic granules after specific interaction with the target cell (see illus.). After target cell binding, the Golgi apparatus, which is an organelle that prepares proteins for secretion, and the granules containing the cytolytic molecules orient toward the target cell in preparation for secretion. Vectorally oriented granule exocytosis then occurs, with the release of cytolytic molecules into the intercellular space between the killer and target cells. This results in the leakage of salts, nucleotides, and proteins from the target cell, leading to cell death. *See* GOLGI APPARATUS.



Lytic cycle. The killer cell binds to the appropriate target cell, triggering the expression of FasL and the reorientation of the Golgi apparatus and the cytolytic granules toward the target cell. The cytolytic granules then fuse with the external membrane, releasing the granule contents, including perforin and granzyme B, into the intercellular space. The granule contents trigger apoptosis and membrane damage. Death receptors, such as Fas, expressed on the target cell are also engaged, leading to apoptosis. After the lytic pathways are initiated in the target cell, the killer cell releases from the target cell and goes on to repeat the entire process.

The cytolytic granules contain a variety of proteins, including a well-characterized protein called perforin and a group of enzymes called serine proteases. It is thought that the perforin can polymerize into a pore-forming complex, perhaps to act as a channel that contributes to the redistribution of small molecules leading to cell death; however, direct evidence that the formation of a pore plays a role in cell-mediated cytotoxicity is lacking. Mice deficient in perforin have been generated, and the killing capacity of killer cells isolated from these mice is severely compromised, as are the immune responses against some pathogens. Although the exact function of this protein remains to be defined, these results demonstrate that perforin is a critical component in the cytolytic mechanism.

Serine proteases, the best characterized being granzyme B, are also released from the cytolytic granules and play an important part in triggering a cellular suicide process referred to as apoptosis or programmed cell death. In apoptosis, a cytotoxic cell triggers an endogenous suicide pathway within the target cell, resulting in nuclear disintegration, DNA fragmentation, and mitochondrial dysfunction. The target cell plays an active role in its own demise, which is consistent with much of the data available on cytotoxic T-lymphocyte-mediated killing.

Upon entering the target cell, granzyme B cleaves other proteases called caspases, triggering a cascade of enzymatic reactions leading to changes in the mitochondria and plasma membrane and DNA fragmentation, ultimately culminating in cell death. It is not exactly understood how granzyme B enters the cell

to initiate the caspase cascade. It was initially thought that perforin formed pores through which the proteases entered to initiate the death cascades; however, a number of observations are inconsistent with this hypothesis, and recent evidence suggests that additional mechanisms for serine protease entry into the cells may exist. Mice made deficient of granzyme B have compromised killer-cell-dependent immune responses, indicating that granzyme B is important for cell-mediated cytotoxicity.

Although it is not clear how the granule contents cause cell death, substantial evidence suggests that these granules play an important role in mediating target cell lysis by killer lymphocytes by both apoptotic and nonapoptotic pathways. Future research efforts will be focused on determining exactly how the granule contents, and specifically granzyme B and perforin, cause cell death in the target cell.

Death receptors. On virtually all cells of the body a number of receptors have been identified that are able to trigger apoptosis when engaged. These receptors are called death receptors. The most studied, and relevant to cytolytic cells, is a receptor called Fas. Fas, when engaged by its ligand (FasL), triggers the caspase cascade leading to the hallmark signs of apoptosis, which include membrane blebbing, chromosomal condensation, nuclear disintegration, DNA fragmentation, and cell death. Thus, the cellular suicide pathway triggered by Fas engagement is similar in some respects to the pathways triggered by granzyme B. If FasL were expressed constitutively on killer cells, this would mean that the killer cell could indiscriminately lyse cells; however, FasL is expressed only after the target cell has been recognized. Since FasL is very rapidly expressed (for example, after the T-cell receptor for antigen is engaged), Fas is triggered only on the target cell that has been identified as a cell destined for death. Fas is expressed on essentially all cells of the body, indicating that all cells can be easily triggered to undergo cell death by the appropriate killer cell. It makes sense for all of the cells in the body to have a built-in self-destruct mechanism so that if a cell becomes pathogenically altered it can be easily triggered to die.

There appears to be no single mechanism that killer cells use to eliminate tumor cells or infected cells. The belief that killer cells employ multiple mechanisms is consistent with the plasticity of the immune system and is important given the wide range of pathogenic organisms that the body must eliminate. Unfortunately, these killer cells may also play a role in certain autoimmune diseases and graft rejection after transplantation. Understanding how these killer cells function and are regulated will be essential for developing therapies to either enhance advantageous or inhibit deleterious cytotoxic lymphocyte responses. *See* IMMUNOLOGY.

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Cytomegalovirus infection

A common asymptomatic infection caused by cytomegalovirus, which can produce life-threatening illnesses in the immature fetus and in immunologically deficient subjects.

The virus. Cytomegalovirus is a member of the herpesvirus group, which asymptomatically infects 50–100% of the normal adult population. Such infections usually take place during the newborn period when the virus can be transmitted from the mother to the baby if the virus is present in the birth canal or in breast milk. Toddlers may also acquire the infection in nurseries. Later in life, the virus may be transmitted by heterosexual or male homosexual activity. After infection, cytomegalovirus remains latent in the body because it cannot be completely eradicated even by a competent immune system. It may be activated and cause illnesses when there is a breakdown of the immune system.

Pathogenesis. Congenital or transplacental cytomegalovirus infection, like asymptomatic infection after birth, is also a fairly common event. With rare exceptions, it too is usually asymptomatic. Congenital cytomegalovirus disease was the first disease recognized to be caused by this virus. It results from transplacental transmission of the virus usually from a mother undergoing initial or primary cytomegalovirus infection during pregnancy. Its manifestations range from subtle sensory neural hearing loss detectable only later in life, to a fulminating multisystem infection and eventual death of the newborn. This important congenital disease occurs in about 1 in 1000 pregnancies.

The only cytomegalovirus illness clearly described in mature, immunologically normal subjects is cytomegalovirus mononucleosis. This is a self-limited illness like infectious mononucleosis, the main manifestation of which is fever. *See* INFECTIOUS MONONUCLEOSIS.

Otherwise, cytomegalovirus illnesses are usually seen only when cellular immunity is deficient. They constitute the most important infection problem after bone marrow and organ transplantations. Manifestations vary from the self-limited cytomegalovirus mononucleosis to more serious organ involvement such as pneumonia, hepatitis, gastrointestinal ulcerations, and widespread dissemination. The virus causing these illnesses may come from activation of the patient's own latent infection, or it may be transmitted from an outside source, usually from latent

cytomegalovirus infecting the graft from a donor. *See* IMMUNOLOGICAL DEFICIENCY; TRANSPLANTATION BIOLOGY.

Cytomegalovirus illnesses are also serious, fairly frequent complications of the acquired immunodeficiency syndrome (AIDS). One reason is that most individuals with human immunodeficiency virus (HIV) infection are already infected with cytomegalovirus. Disease manifestations are similar to what is seen in transplant cases, except they may be more severe. Cytomegalovirus retinitis is a typical problem associated with advanced AIDS. Without treatment, the retina is progressively destroyed such that blindness of one or both eyes is inevitable. *See* ACQUIRED IMMUNE DEFICIENCY SYNDROME (AIDS).

Treatment. Cytomegalovirus diseases can be treated with two antivirals, ganciclovir or foscarnet, with varying degrees of success. Cytomegalovirus pneumonia in the bone marrow transplant recipient cannot be cured by antivirals alone because it probably has an immunopathologic component. Cytomegalovirus diseases in persons with AIDS can be contained but not cured by specific treatment. For example, ganciclovir treatment of cytomegalovirus retinitis is effective only as long as maintenance therapy is continued. *See* ANIMAL VIRUS. Monto Ho

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Cytoplasm

That portion of living cells bordered externally by the plasma membrane (cell membrane) and internally by the nuclear envelope. In the terminology of classical cytology, the substance in living cells and in living organisms not compartmentalized into cells was called protoplasm. It was assumed at the time that the protoplasm of various cells was similar in structure and chemistry. Results of research on cell chemistry and ultrastructure after about 1960 showed that each cell type had a recognizably different "protoplasm." Primarily for that reason, the term protoplasm gradually fell into disuse in contemporary biology. The terms cytoplasm and nucleoplasm have been retained and are used descriptively; they are used almost synonymously with the terms cytosome (body of cytoplasm) and nucleus, respectively.

Observation by optical microscopy. Optical microscopes are limited in resolving power at the level of about 0.2 micrometer. Therefore only the larger structures in the cell could be observed and named by early cytologists. Structures visible in the light microscope include the nucleus, with its chromosomes and nucleolus, and the cytoplasm, with its larger organelles such as chloroplasts (in plants only), mitochondria, Golgi bodies, and centrosome (containing two centrioles), as well as inclusions such as lipid droplets, pigment granules, yolk (in eggs), starch grains, and vacuoles. The main impression that early microscopists gained from observations of cells with an ordinary bright-field microscope was that cells were transparent and replete with particles of different types, and that many exhibited motion of some kind, usually Brownian motion (thermally induced collisions between molecules and granules in the cell).

Modern optical microscopes operate in modes that produce vastly improved contrast. Dark-field and phase-contrast microscopy greatly increases the visibility of cytoplasmic particles of all kinds. Interference microscopes show the shapes of cells by representing differences in optical path (refractive index difference times the thickness) in brightness contrast or color contrast. Differential interference microscopes render all cytoplasmic surfaces shadowcast, and provide a more detailed view of the various interfaces in cells caused by the existence of organelles and inclusions, vacuolar membrane systems, and various kinds of filamentous structures. Polarizing microscopes reveal ordered or crystalline structures (such as the mitotic spindle, stress fibers, and myofibrils) when they are doubly refracting, or birefringent. Fluorescence microscopes reveal naturally fluorescent substances (fluorochromes) such as chlorophyll in plants, or fluorchrome-labeled antibodies to specific protein molecules. These optical probes have made it possible to discover the location of enzymes, hormones, and structural proteins in the cytoplasm, sometimes even while cells are alive. *See* MICROSCOPE.

Regional differentiation. Many cells, especially the single-celled organisms or protistans, have regional cytoplasmic differentiation. The outer region is the cortex or ectoplasm, and the inner region is the endoplasm. In some echinoderm eggs, for example, endoplasmic organelles and inclusions can be sedimented at moderate centrifugal accelerations, while cortical granules that are denser than most other particles remain fixed in the cortex. In many cases the cortical layer is a gel made up of a meshwork of cytoskeletal fibers, especially F-actin. In many ciliates the cortical layer contains organelles peculiar to that region. For example, the cortex of *Paramecium* contains the basal bodies of cilia and a system of fibrils associated with them, as well as rows of trichocysts which are fired like tiny javelins when the *Paramecium* is injured. In amebae, movement requires that the cell maintain an outer ectoplasmic gel surrounding the more fluid flowing endoplasm, which nevertheless also has some gellike properties.

Apparently the cytoplasm of nearly all cells is contractile when stimulated. In some cells the machinery of contraction is localized in the cortex (for example, the ciliates *Stentor, Spirastomum*, and *Vorticella;* the flagellates *Euglena* and *Peranema;* the testacean *Difflugma;* and many sporozoans).

Centrifugal sedimentation of cytoplasm. The organelles and inclusions of some cells can be caused to sediment into layers based on their density by the application of centrifugal acceleration. Particles with high density are sedimented to the centrifugal pole, while less dense (buoyant) particles are moved to the centripetal pole. In classical experiments by

E. N. Harvey and E. B. Harvey in the 1930s, it was shown that eggs of the sea urchin *Arbacia* could be stratified while under observation on a cushion of sucrose solution in a special microscope-centrifuge, actually a centrifuge with the microscope built into it. If the eggs were centrifuged while in a sucrose density gradient, they were pulled into two halves by the force caused by the centrifugal acceleration acting upon the density difference between the upper and lower halves of the egg and the local density in the gradient.

Centrifugation experiments clearly showed that organelles and inclusions fall into classes differentiated by their density, and focused attention on the properties of the ground cytoplasm (or hyaloplasm) in which the organelles and inclusions were bathed. The early centrifugation experiments led to later efforts directed at measuring the consistency of the ground cytoplasm, and to isolation and purification of fractions of cytoplasmic particles. *See* CENTRIFUGATION.

Observation by electron microscopes. The fine structure of cells could not be studied until the development of electron microscopes and a technique for cutting very thin sections (slices) of cells embedded in a suitable plastic. The first ultramicrotome for slicing cells embedded in plastic was developed by K. R. Porter. The scanning electron microscope (SEM) permits a detailed view of the cell surface or into a cell that has been opened or rendered partially transparent by treatment with electrically conductive fixative compounds. The resolution of the SEM is limited at between 2 and 5 nanometers, depending on instrument design. *See* SCANNING ELECTRON MICROSCOPE.

The conventional transmission electron microscope (CTEM or TEM) is used to produce images and electron micrographs of thin sections (about 40–60 nm thick). The resolving power of the TEM is about 0.2 nm, about a thousand times greater than any light microscope.

The high-voltage electron microscope (HVEM) is similar in design to the TEM but operates with an electron beam of $1–3 \times 10^6$ eV, about 10 times greater voltage than the CTEM. The higher penetrating power and reduced specimen damage at higher voltages make it possible to obtain stereo images of thick sections or whole cells up to 2–5 micrometers in thickness. The HVEM has been very important in obtaining a comprehensive view of the relationships of parts of cells not appreciated from thin sections. *See* ELECTRON MICROSCOPE.

Cytoplasmic ultrastructure. Electron microscopy revolutionized concepts of cell structure and function. Organelles too small to be seen clearly with light microscopes were discovered, named, and characterized after 1955. Organelles previously known were characterized as to their detailed inner morphology.

Centrioles. The two centrioles within the centrosome or central body of the cytoplasm appear as barely visible spherical dots in the optical microscope. In the electron microscope the two centrioles can be seen to be short bundles of 27 tubules arranged in a characteristic pattern, each bundle perpendicular to the other. Centrioles are present in most animal cells and absent in most plant cells, and are organizing centers for the microtubular structures in the mitotic spindles of dividing animal cells and in cilia and flagella. The basal bodies of the latter are related to centrioles. *See* CENTRIOLE.

Endoplasmic reticulum. In cells known to synthesize proteins for secretion (for example, cells of the pancreas), the cytoplasm just outside the nucleus was known from optical microscopy to be basophilic. This region of the cell, called the ergastoplasm, has an affinity for basic stains long known to mark the presence of nucleic acids. In the electron microscope, this region of the cell was found to contain massive amounts of rough endoplasmic reticulum consisting of stacks of flat membranous cisternae covered with ribosomes, particles of ribonucleoprotein which play an important role in protein synthesis. Face-on views of the endoplasmic reticulum show the ribosomes in lines connected by delicate strands of messenger ribonucleic acid, macromolecules that serve as an intermediate carrier of genetic information for the synthesis of specific protein molecules. The rough endoplasmic reticulum has been shown to be a temporary repository for proteins to be exported from the cell or incorporated into membranes. Smooth endoplasmic reticulum is more likely to be tubular than flat, and lacks attached ribosomes; it is involved in the synthesis and storage of other macromolecules. *See* ENDOPLASMIC RETICULUM; RIBOSOMES.

Golgi complex. The Golgi complex is another class of organelles consisting of stacks of membranous sacs or cisternae that are slightly curving in shape. The Golgi complex serves to package proteins that the cell will secrete. The edges of the Golgi cisternae break off and form Golgi vesicles containing the secretions. *See* GOLGI APPARATUS.

Mitochondria. Mitochondria, usually rod-shaped particles 0.5–3 μm long, consist of two membranes, the inner one thrown into folds called christae, on which are located assemblies of respiratory pigments and enzymes. Mitochondria are often spoken of as the powerhouses of the cell. Their matrix surrounding the inner membrane folds contains the enzymes that catalyze the oxidative phase of glycolysis (carbohydrate breakdown). Mitochondria each contain a circular strand of deoxyribonucleic acid (DNA), which contains the genetic information coding for some, but not all, of the mitochondrial enzymes and structural proteins. *See* MITOCHONDRIA.

Lysosomes. Lysosomes are cytoplasmic particles in the same size range as mitochondria, and can be recognized by an electron-dense matrix with no folded inner membrane. Lysosomes differ in density from mitochondria, and thus can be separated from them by density gradient centrifugation. Lysosomes contain several hydrolytic enzymes that can catalyze the breakdown of various classes of macromolecules. When cells ingest foreign material (endocytosis), lysosomes fuse with the vacuole containing the ingested material and supply enzymes required for its

intracellular digestion. *See* ENDOCYTOSIS; LYSOSOME.

Microbodies. Another class of cytoplasmic particles are the microbodies, consisting of peroxisomes and glyoxisomes, particles that contain different groups of oxidative enzymes.

Chemical deposits. Cells contain inclusions that are not functional entities but, rather, deposits of certain chemicals. Many cells contain small droplets of fat or oil that are buoyant and can be displaced by centrifugation to the centripetal pole of the cell. Certain large amebas contain bipyrimidal or flat crystals of triuret, a nitrogen excretion product of purine and pyrimidine metabolism. In the same cells are many so-called refractile bodies containing calcium and magnesium polyphosphate compounds. Plant cells contain starch particles which can be broken down to glucose to obtain chemical energy; many animal cells contain glycogen deposits that play a similar role. The inclusions found in various cell types are more numerous and varied than the organelles. They can be removed by centrifugation and microsurgery without harming the cell, and are therefore nonessential to life.

Role in cell division. When the cell divides, a new organelle, the mitotic spindle, is formed, which divides the replicated chromosomes into equal groupings for the daughter nuclei to be formed. In some cells the mitotic spindle forms, at least in part, while the chromosomes are still within the nuclear envelope. In other cells the spindle itself is entirely intranuclear.

Cytoplasmic division (cytokinesis) in animal cells and many protists is accomplished by a contractile ring of microfilaments composed of the protein F-actin. Contraction involves the interaction of F-actin with another protein, myosin. In plant cells cytokinesis is accomplished by the transport of Golgi vesicles containing precursors of the new cell plate along microtubules of the phragmoplast, a reorganized remnant of the mitotic spindle. *See* CELL DIVISION; CYTOKINE.

Cytochemistry. Cytoplasm contains mostly water, from 80 to 97% in different cells, except for spores and other inactive forms of living material, in which water may be present in lesser amounts. The dry mass of cells consists mainly of macromolecules: proteins, carbohydrates, nucleic acids, and lipids associated with membranes. The small molecules present in cells are mainly metabolites or metabolic intermediates. The principal ions other than the hydrogen and hydroxyl ions of water are the cations of potassium, sodium, magnesium, and calcium, and the anions chloride and bicarbonate. The concentrations of these ions are maintained within narrow physiological limits by ion-pumping mechanisms and the presence of natural chelators or ion-binding substances. Therefore analyses of whole cells for calcium ion concentration often give values two to three orders of magnitude higher than the concentration of free calcium ions in cytoplasm.

Many other elements are present in cytoplasm in smaller amounts. Iron is found in cytochrome pigments in mitochondria; magnesium is present in chlorophyll in chloroplasts; copper, zinc, iodine, bromine, and several other elements are present in trace quantities.

The presence of certain chemical compounds in the cell can be measured quantitatively by spectroscopic techniques applied either to living cells or to fixed cells subjected to chemical reactions in place. Cytochrome pigments, for example, can be demonstrated in living mitochondria by their absorption bands in the blue region of the spectrum. The same holds for different forms of chlorophyll and accessory photosynthetic pigments in plant cells. Nucleic acid absorption can be detected in the ultraviolet region of the spectrum.

By carrying out chemical reactions in fixed cells, it is possible to localize proteins, carbohydrates, and the two basic types of nucleic acids, DNA and ribonucleic acid (RNA). Enzymes can be localized by chemical reactions that demonstrate the product of the reactions they catalyze.

One of the most powerful techniques in cytochemistry is cell fractionation. Cells are first homogenized or otherwise disrupted in a medium that keeps organelles and inclusions intact and, if possible, functional. The resulting suspension of cell organelles and inclusions is centrifuged first at low accelerations to remove nuclei and other heavy particles, and then the homogenate is centrifuged at a higher acceleration in a manner designed to separate particle classes on the basis of density. In this way it is possible to obtain a nuclear fraction, a mitochondrial fraction, and so on. The purity of each fraction is monitored by optical or electron microscopy, and when the desired criterion of purity has been met, fractions are submitted to chemical analysis. In this way it has been possible to construct lists of substances, especially enzymes, known to be present in different organelles or cell compartments.

The many proteins that lack enzymatic function may be regulatory or structural in function. The best way to identify and localize these is by fluorescent-antibody techniques. A given protein is first isolated and purified. Repeated injections of the purified protein (an antigen) into rabbits causes the production of a rabbit antibody to that protein. Antibody isolated from the blood serum of the rabbit and subsequently purified is then used as a probe to look for the antigenic protein by using fluorescent antibodies to rabbit serum proteins produced by some other animal, such as a goat or sheep. This technique is known as the indirect immunofluorescent technique or the "sandwich" technique. *See* CYTOCHEMISTRY; IMMUNOFLUORESCENCE.

Cytoplasmic rheology. In the first two decades of this century a controversy arose as to whether cytoplasm was fluid or gel. Efforts to resolve the controversy by objective physical experiments were not at first successful, because there was insufficient knowledge of the complexity of cytoplasmic organization, and because certain "simplifying" assumptions turned out to be misleading.

Fluids in which the rate of shear (flow) is proportional to the applied shear stress (force) are

called newtonian fluids. Therefore fluids with a nonlinear force-flow relationship are designated nonnewtonian. The slope of the force-flow relationship for a fluid is defined as its fluidity, while the reciprocal of the slope is viscosity, a measure of fluid resistance. Fluids with nonlinear force-flow relations obviously exhibit different slopes and therefore different viscosities at different rates of shear; that is, they exhibit a different fluid resistance at different rates of flow. Thixotropy is a reduction in viscosity at higher flow rates, and indicates a breakage of affinities or bonds contributing to fluid resistance. Dilatancy is the opposite effect: an increase in resistance with increased rate of flow. *See* BIORHEOLOGY.

The first objective methods used in an attempt to gain quantitative information about the consistency of cytoplasm relied on the assumption, which seemed reasonable at the time, that cytoplasm might be a newtonian fluid. Since this assumption disregarded the possibility that cytoplasm might be structured, the fact that the methods used applied forces that destroyed cytoplasmic structure was overlooked.

It is now clear that cytoplasm is highly nonlinear in rheological behavior due to the presence of cytoskeletal proteins in the form of microfilaments, microtubules, intermediate filaments, and various other cytoskeletal proteins in dynamic association with one another. These skeletal elements can confer sufficient structure on cytoplasm in some cells so that a measurable yield force is required to cause any flow at all. The presence of a yield point classifies any fluid as a gel. Some gels can be strong, as in the case of gelatin or agar, while other cytoplasmic gels are weak in comparison. Even weak gel structure is important, because the forces acting upon the cytoplasm are also weak. Cytoplasm in some cells also shows the property of viscoelasticity. Viscoelastic fluids have a "memory" for their original shape, and when deformed tend to recoil toward their original shape.

It is very difficult to characterize the rheological properties of cytoplasm quantitatively, because they depend on the force applied in order to make the measurement. Reliable viscosity values have been found for cytoplasm in the range of 5 centipoise to 10 poise or 0.005 to 1.0 pascal · second (water has a viscosity of about 1 centipoise or 0.001 Pa · s).

In some cells it has been possible to differentiate between the macroviscosity, the viscosity of the cytosol and cytoskeleton, and the microviscosity, the viscosity of the cytosol alone. The former is always an order of magnitude (or more) more viscous than the latter.

Important experiments carried out by Y. Hiramoto showed that the yield point and viscoelastic constants of cytoplasm could be measured by the displacement of an iron sphere inside a cell by an electromagnet. These studies showed the complex nature of the rheological properties of cytoplasm, and the fact that these properties can change with the life cycle of the cell.

Ground cytoplasm: cytosol and cytoskeleton. Sedimentation of cells by centrifugation shows that organelles and inclusions can be separated from the ground cytoplasm, the fluid phase of the cytoplasm in which they are suspended. The ground cytoplasm in turn has been shown to consist of a cytoskeletal network and the cytosol, the fluid in which the cytoskeleton is bathed.

The cytoskeleton consists of several biopolymers of wide distribution in cells. Microtubules have been observed in electron micrographs of a vast number of different cell types. They are about 25 nm in diameter and very long. They consist of the protein tubulin, and are frequently covered by a fuzzy layer of microtubule-associated proteins. Microfilaments of more than one type have been observed. The most common and ubiquitous are those composed of F-actin, one of the most common proteins in muscle of all kinds. Microfilaments are usually 6–7 nm in diameter and several micrometers long. If composed of F-actin, they can be decorated in a characteristic arrowhead pattern by the S-1 subfragment of skeletal muscle myosin to reveal their polarity. Intermediate filaments of several types are known. Most have a diameter of about 10 nm, and are sometimes called 10-nm filaments. At least some of these intermediate filaments are related to the protein keratin, which is present in skin, nails, hair, and horns.

Many proteins have been found that appear to have either a regulatory or cross-linking (gel-forming) function. They bind to one or more of the cytoskeletal proteins and act to control contractility, motility, or cytoplasmic consistency. The patterns of distribution of cytoskeletal proteins have been demonstrated by fluorescent antibody probes which show the presence of a given specific protein by labeling it with a fluorescent marker using the specificity of an antigen-antibody reaction. *See* CYTOSKELETON.

Cytoplasmic motility. In most cells the smaller particles exhibit brownian motion due to thermal agitation. In some cells lacking extensive cytoskeletal structure, particles can be moved freely around the cell by brownian motion. In others they are restricted by the cytoskeletal elements. *See* BROWNIAN MOVEMENT.

Particles of various types may also undergo saltatory motions which carry them farther than brownian motion possibly could. Such excursions are usually either linear or curvilinear, and result from the interaction of a particle with an element of the cytoskeleton such as one or more microtubules or microfilaments. Single particles or trains of particles may exhibit saltation, which is a jumping motion.

When masses of particles move together, they are usually accompanied by a flow of the cytoplasm around them called cytoplasmic streaming. Streaming of cytoplasm is best seen in cells of *Nitessa* or *Chara* (green algae), in amebas and paramecia (protists), in cellular slime molds (fungi), and in many cells (for example, eggs and leukocytes) of higher animals. There is evidence to indicate that cytoplasmic streaming involves the interaction of the proteins F-actin (microfilaments) and myosin. However, the manner of interaction and the cellular regulatory

systems appear to differ widely. The movement of cells of both protists and higher organisms involves contractile, cytoskeletal, and regulatory proteins which exhibit differences from species to species and from cell to cell. *See* CELL MOTILITY.

Perhaps the most intricate and important form of motility is that which occurs in neurons (nerve cells) of all types: axoplasmic transport. Neurons synthesize transmitter substances near their nuclei in the cell body. Vesicles containing these transmitter substances must be transported long distances (up to 6 ft or 2 m in humans) in order to be released at the synapses where one nerve cell excites the next. Anterograde axoplasmic transport brings the vesicles to the nerve endings, whereas retrograde transport carries chemical messages back to the nerve cell body. Failure of this transport mechanism in certain diseases (for example, Alzheimer's disease and senility of old age) is due to the failure of this important process. *See* ALZHEIMER'S DISEASE.

Muscular movement is made possible by virtue of paracrystalline interdigitating arrays of F-actin thin filaments and myosin thick filaments in the myofibrils in the cytoplasm of muscle fibers. Muscle contraction is a highly specialized form of contractility, which is one of the basic properties of cytoplasm. *See* CELL BIOLOGY; MUSCLE; MUSCLE PROTEINS.

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Cytoskeleton

A system of filaments found in the cytoplasm of cells which is responsible for cell shape, cell locomotion and elasticity, interconnection of the major cytoplasmic organelles, cell division, chromosome organization and movement, and the adhesion of a cell to a surface or to other cells.

Since this filamentous system is too small to be resolved by light microscopy, its several subcomponents, structural features, and cytoplasmic associations were not discovered until the electron microscope was developed. With this instrument, which is capable of magnifying a section of a cell many thousands of times, three major classes of filaments could be resolved on the basis of their diameter and cytoplasmic distribution: actin filaments (or microfilaments) each with an average diameter of 6 nanometers, microtubules with an average diameter of 25 nm, and intermediate filaments whose diameter of 10 nm is intermediate to that of the other two classes. In muscle cells, an additional class of thick filaments, known as myosin filaments, is found whose function is to interact with actin filaments and generate the force necessary for muscle contraction. The presence of this system of filaments in all cells, as well as their diversity in structure and cytoplasmic distribution, has been recognized only in the modern period of biology.

A technique that has greatly facilitated the visualization of these filaments as well as the analysis of their chemical composition is immunofluorescence applied to cells grown in tissue culture. In this technique, the protein that is to be localized is injected into an animal to elicit an antibody response. The resulting antibody is highly specific for the protein. Following several purification steps, the antibody is chemically attached to a fluorescent dye. The cells, which are usually grown on a glass cover slip, are fixed with a cross-linking agent such as formaldehyde in order to stabilize their cytoskeletons. They are subsequently rendered permeable to the fluorescent antibody by exposure to a detergent or to a polar solvent such as ethanol or acetone. The cells are then treated with the fluorescent antibody solution for a few minutes; unbound antibodies are washed away, and the cover slip is examined under a regular light microscope equipped with fluorescence optics. The fluorescent antibody binds to the protein it is specific for, and upon excitation of the fluorescent dye, the location of that protein is revealed. Extensions of this technique involve the use of antibodies and electron microscopy. In this technique, known as immunoelectron microscopy, the resolution of subcellular localization is increased appreciably. Furthermore, it can be used to detect the binding site of one protein to another, thereby expanding the detail at which cytoskeletal molecular interactions among each other and in the cytoplasm can be understood. *See* IMMUNOFLUORESCENCE.

Actin filaments. Actin, a protein highly conserved in evolution, is the main structural component of actin filaments in all cell types, both muscle and nonmuscle. As revealed by immunofluorescence and electron microscopy, actin filaments assume a variety of configurations depending on the type of cell and the state it is in. They extend a considerable distance through the cytoplasm in the form of bundles (**Fig. 1***a*). These bundles are also known as stress fibers since they are important in determining the elongated shape of the cell and in enabling the cell to adhere to the substrate and spread out on it. In a single migratory or resting cell, most of the actin filament bundles are located just below the cell membrane and are frequently in close association with it. The sites where the bundles make contact with the cell membrane, known as adhesion plaques, correspond to the sites of cell-substratum contact. It has been estimated from time-lapse motion pictures that most of the actin filament bundles are oriented parallel to the long axis of the cell and shift their orientation as the cell changes direction; the bundles reorient themselves over a period of 15 to 60 min.

Actin filaments can exist in forms other than straight bundles. In rounded cells that do not adhere strongly to the substrate (such as dividing cells and cancer cells), the filaments form an amorphous meshwork that is quite distinct from the highly organized bundles. This filamentous meshwork is usually located in the cell cortex, which is the area just beneath the cell membrane. It is also found in areas of membrane ruffling in motile cells.

Membrane ruffles are sheetlike extensions of the cytoplasm which function as the main locomotory organ of the cells (**Fig. 2**). The membrane ruffle forms transient contacts with the substrate on which the cell rests, stretching the rest of the cell among several temporary adhesions. When the cell is spread out to its fullest extent, it maintains one or two primary ruffles, each of which tends to lead the cell in a particular direction. When the transient contacts of a membrane ruffle are converted to permanent contacts, they become adhesion plaques and function to organize long actin filament bundles. Thus, the two filamentous states, actin filament bundles and actin filament meshworks, are interconvertible polymeric states of the same molecule. They appear to be the extremes of a gradient, and the ratio between them depends on a variety of factors, including the extent of cell motility, the age of the cell, the degree of adhesion to the substrate, and the stage of the cell cycle. Bundles give the cell its tensile strength, adhesive capability, and structural support, while meshworks provide elastic support and force for cell locomotion.

Actin filaments are involved in two main cellular functions: cell motility, and cell shape and elasticity (Fig. 1*a*). These broad types of function are mediated by specific accessory proteins that interact with actin to change the properties, functional associations, or the polymeric state of actin filaments. Thus, the cell has molecules which regulate the polymerization of actin to single filaments, molecules which regulate the assembly of single actin filaments to filament bundles, molecules which interconvert actin filament bundles and actin filament meshworks, molecules which mediate the interaction of actin filaments with the plasma membrane, and finally molecules, such as myosin, which enable actin filaments to move the cell as a whole or mediate specific motile activities to the cytoplasm, such as cell cleavage. In turn, the spatial and temporal interaction of these proteins with actin is regulated, in part, by the cytoplasmic hydrogen ion concentration (pH) and the calcium ion concentration. In the cytoplasm these regulatory mechanisms coexist in a dynamic equilibrium,

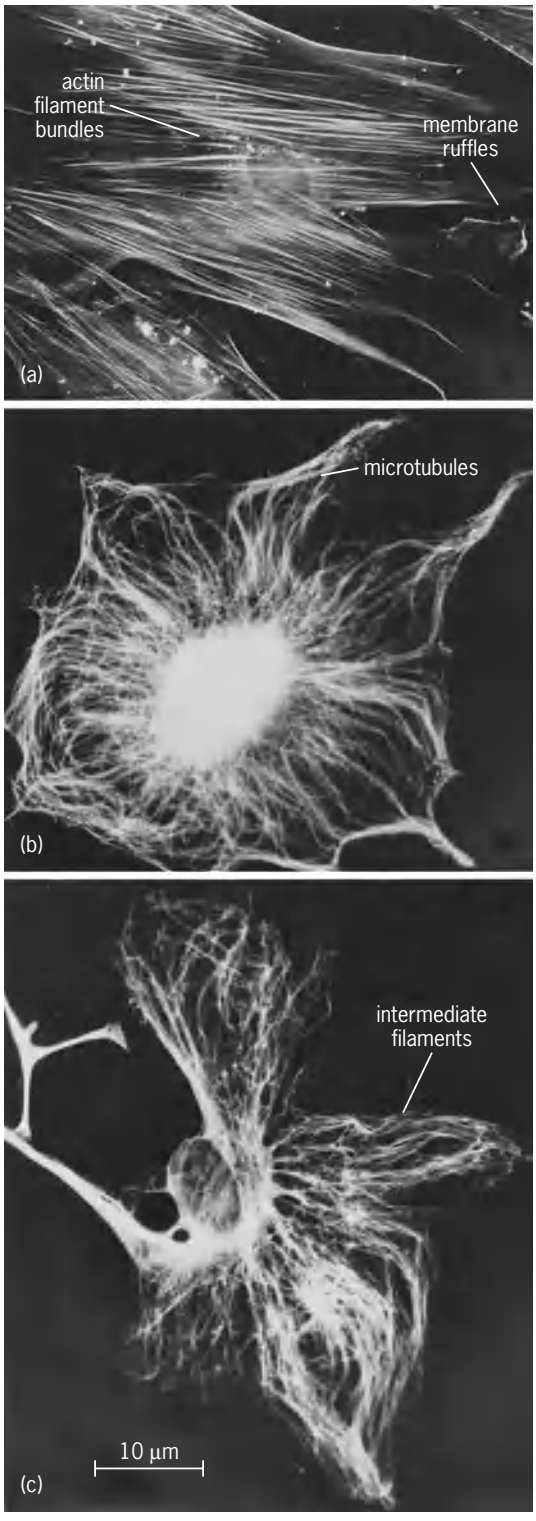

Fig. 1. Immunofluorescence with human cells grown in tissue culture. (*a*) Distribution of actin in filament bundles and membrane ruffles using fluorescent antibodies to muscle actin. (*b*) Distribution of microtubules using fluorescent antibodies to their major subunit tubulin. (*c*) Distribution of intermediate filaments using fluorescent antibodies to their major subunit from fibroblasts.

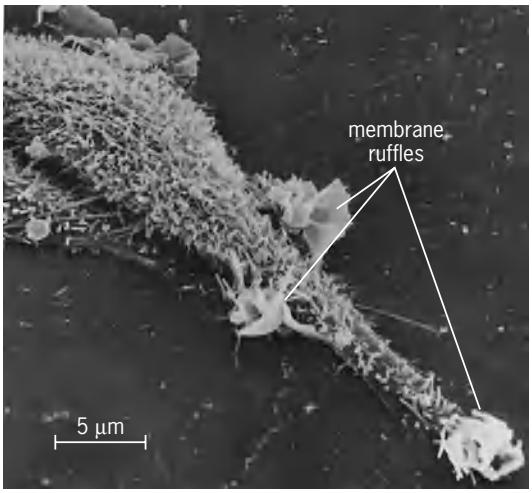

Fig. 2. Scanning electron micrograph of the surface of a human fibroblast grown in tissue culture. The small protrusions on the cell surface are known as microvilli or microspikes and are composed of a core of 20–30 actin filaments.

many of them operating simultaneously, to govern cell motility and maintenance of cytoplasmic structure.

The importance of accessory actin-binding proteins in the maintenance of cell shape and elasticity has been demonstrated through the study of certain severe human hemolytic anemias, such as hereditary spherocytosis, pyropoikilolytosis, and elliptocytosis. In these diseases one or more of the actin-binding proteins are defective. Normally, this battery of proteins binds to and crosslinks actin filaments, forming a pliable and deformable network under the plasma membrane of the red blood cell. Mutations in any one of these accessory proteins cause structural alterations in this skeletal network, resulting in the fragmentation of the plasma membrane as the cells deform their shapes during circulation. The consequence is cell lysis and hemolysis. A number of these accessory proteins are now known to be common to red blood cells, lens cells, muscle cells, and nerve cells, among others, in association with actin.

A main function of actin filaments is to interact with myosin filaments to generate the force that allows muscles to contract, and many nonmuscle cells to move. The interaction of these two filament systems is extremely well coordinated by a different set of proteins that interact with either actin filaments or myosin filaments. The interaction of these proteins with their respective filaments is regulated by Ca^{2+} ions, by the H^{+} ion concentration in the cytoplasm, and by direct covalent modifications such as phosphorylation. *See* MUSCLE PROTEINS.

Microtubules. These slender cylindrical structures exhibit a cytoplasmic distribution distinct from actin filaments. Microtubules originate in structures that are closely associated with the outside surface of the nucleus known as centrioles. Fluorescent antibodies specific to the major structural protein of these filaments, known as tubulin, reveal a wavy system of filaments originating in the centriole area and extending to the cell's outer membrane (Fig. 1*b*). Unlike the other two classes of filaments, microtubules are highly unstable structures and appear to be in a constant state of polymerization-depolymerization. During mitosis, they rapidly disaggregate and reorganize around the mitotic spindle, where they function in moving the chromosomes to the two newly forming daughter cells. In the resting cell, microtubules are involved in the determination of a cell's shape and in cytoplasmic transport of vesicles and other small organelles. In conjunction with the centrioles, they also function as the "information processing center" of the cell since they translate external stimuli into decisions about the direction of locomotion of a cell. In more specialized cells, like the sperm cell, they function to propagate the cell through body fluids. *See* CENTRIOLE; MITOSIS.

Intermediate filaments. These filaments function as the true cytoskeleton. Unlike microtubules and actin filaments, intermediate filaments are very stable structures. If a cell is exposed to a mild nonionic detergent, most of the microtubules, actin filaments, and various cytoplasmic organelles are rapidly removed, but intermediate filaments and nuclei are left insoluble. In the intact cell, they anchor the nucleus, positioning it within the cytoplasmic space. During mitosis, they form a filamentous cage around the mitotic spindle in a fixed place during chromosome movement. As seen using immunofluorescence (Fig. 1*c*), intermediate filaments have a cytoplasmic distribution independent of actin filaments and microtubules. However, their distribution in the cytoplasm depends very much on the type of cell and its stage in differentiation. For example, in undiffferentiated muscle cells these filaments are found scattered in the cytoplasm, but in mature, fully differentiated cells they wrap around the individual contractile units, the myofibrils, and function to interconnect them to the plasma membrane. In chicken red blood cells, which contain a nucleus, these filaments function to anchor the centrally located nucleus to the plasma membrane. In mammalian red blood cells, however, which lose their nuclei, these filaments are removed prior to enucleation. In all epithelial cells, and epidermal cells in particular, a special class of intermediate filaments, known as keratin filaments, interconnect the nuclei to special areas of cell-cell contact at the plasma membrane known as desmosomes. In this fashion, an epithelial layer is integreted mechanically both intercellularly and intracellularly to withstand tension and possess elastic properties. Therefore, intermediate filaments are a fundamental component of the morphogenesis of certain differentiating cells. *See* MUSCLE.

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Dacite

An igneous volcanic rock that is the extrusive equivalent of granodiorite, according to the IUGS (International Union of Geological Sciences) classification scheme. The overall color of dacite lava is generally light to medium gray but can appear pinkish or tan when oxidized or weathered. The color index of dacite averages 12–14%, denoting that mafic or dark minerals (hornblende, biotite, hypersthene, augite, titanomagnetite, and ilmenite) make up 12–14% of the rock. As an extrusive igneous rock, dacite is usually porphyritic in texture, meaning it contains 0–50% phenocrysts (visible crystals), in a fine-grained or glassy matrix called the groundmass (**Fig. 1**). The phenocryst minerals typically include some combination of plagioclase, hypersthene, augite, hornblende, biotite, quartz, potassium feldspar, titanomagnetite, and ilmenite, depending on chemical composition and equilibration conditions (pressure, temperature, and oxygen fugacity) of the magma prior to eruption. Dacite may also occur as aphyric (no visible crystals) lava, composed entirely of microcrystalline matrix minerals or glass (vitrophyric). *See* GRANODIORITE; IGNEOUS ROCKS; LAVA; MAGMA; PHENOCRYST.

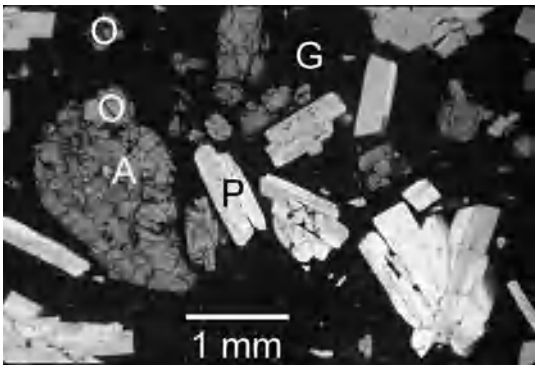

Fig. 1. Photomicrograph of a dacite from the Zitacuaro–Valle de Bravo region of central Mexico. P = plagioclase, O = orthopyroxene (hypersthene), A = augite, G = glassy groundmass.

Fig. 2. Dacite domes from the Zitacuaro volcanic complex of central Mexico. (*Photograph by D. Blatter*)

Dacite is intermediate between andesite and rhyolite in the basalt-andesite-dacite-rhyolite continuum and is classified as having 63–70 wt % silica (SiO_2). Rocks within the dacite SiO_2 range that contain greater than 2.4 wt % potassium oxide (K_2O) are known as latites, and rocks with total alkalis ($Na_2O + K_2O$) greater than ~7 wt % are known as trachytes. Dacite alumina (Al_2O_3) concentrations average about 16 wt %. *See* ANDESITE; LATITE; RHYOLITE; TRACHYTE.

Dacite magma is generated from basaltic or andesitic magmas through magma mixing, assimilation of crustal material, or fractional crystallization. These processes are most common at convergent margins where dacite eruptions have recently occurred, such as at the volcanoes Mount Pinatubo and Mount St. Helens. Because of the high SiO_2 contents and high crystallinities (up to 50% visible crystals) of many dacites, this lava type erupts to form steep-sided domes (**Fig. 2**) and is often associated with explosive eruptions, which form pyroclastic and air-fall deposits. *See* BASALT; PYROCLASTIC ROCKS; VOLCANO; VOLCANOLOGY.

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Dairy cattle production

A section of animal agriculture concerned with the production of milk and milk products. The mammary gland in mammals was developed for the nourishment of their young. Over the centuries humans have selected some mammals, especially the cow, goat, and water buffalo, for their ability to produce quantities of milk in excess of the requirements of their young. At the same time humans have improved the management and nutrition of the animals so that they can produce milk more efficiently. The production of milk includes selection of animals, breeding, raising young stock, feed production, nutrition, housing, milking, milk handling, sanitation, disease control, and disposal of milk, surplus animals, and manure.

The cattle industry dates back many centuries. In Libya, Africa, friezes dating back to 9000 B.C. show domesticated cows. Written records as far back as 6000 B.C. of Old Mesopotamia (now Iraq) indicate that dairying was very highly developed at that time. These people were the first to make butter. Egyptian records show that milk, butter, and cheese were used as early as 3000 B.C.

Dairy development in America. Dairying in the United States had its principal origin from northern Europe. The first cattle were brought to the West Indies by Columbus in 1495. The first importations to the United States were made to Jamestown, Virginia, in 1611, and the Plymouth Colony in 1624.

In the United States up to 1850 most families kept a cow and the milk and its products were consumed by the family. Excess milk was used for cheese and butter, but little was sold since transportation was by horse and wagon and no refrigeration was available.

After 1850 dairying became an industry as a result of several major factors. In 1841 the first regular shipment of milk by rail was made from Orange County, New York, to New York City. The first cheese factory was built in 1851 in Oneida, New York. In 1856 Gail Borden invented the condensed milk process. The first butter factory was started in New York State in 1856, and the first creamery was established in Iowa in 1871. The separator was invented by Gustav DeLaval in 1878, and in 1890 the Babcock test (to determine fat content) was introduced, which allowed for the impartial buying of milk. The first pasteurizing machine was introduced in 1895, and the first compulsory pasteurization law was passed in Chicago in 1908. Refrigeration came into use in 1880. Contributing to the development of the dairy industry was the establishment of experiment stations, which conducted research on feeding, breeding, and management of cattle. This was accompanied by the organization of the agricultural extension service and other educational facilities to carry information to the dairy farmers. Artificial insemination was introduced in 1938 and is now an important segment of the dairy cattle industry.

World dairy production. The leading milk-producing countries in the world, in decreasing order of annual milk production, are Russia, United States, France, Germany, United Kingdom, Netherlands, Italy, and Canada. Figures are not available for most of the Asian countries, which include the mainland of China and India. Most of the leading milk-producing countries are located in the temperate regions. Dairy cattle numbers and milk production are somewhat more limited in the tropical and semitropical regions. Some countries, namely, New Zealand and Denmark, are not listed among the 10 top countries for milk production but are major export countries because of their large per capita production in comparison to consumption.

Dairy production in United States. The five leading dairy states are Wisconsin, California, New York, Minnesota, and Pennsylvania. The major milk-producing areas are the Midwest and Northeast.

Even though there has been a marked decrease in the number of milk cows since 1940, the annual milk production has increased because of the marked increase in production per cow. A decrease in the number of dairy farms producing milk has been accompanied by an increase in the number of cows per farm and in the number of cows handled by each dairy farmer. Smaller inefficient producers have generally been forced out of the dairy industry, and only the larger, more specialized farms remain.

Milk is used for many products. Approximately 38% goes into fluid milk and cream. Of the remaining milk, about 19% is used for butter, 28% for cheese, and 15% for other products, with frozen dairy products being the most important. *See* BUTTER; CHEESE; ICE CREAM; MILK.

Dairy breeds. Most of the milk produced in the United States is by five major dairy breeds: Ayrshire, Brown Swiss, Guernsey, Holstein-Friesian, and Jersey. Approximately 12% of the dairy cattle are registered purebreds. The first breed registry association was organized by the Jersey dairy producers in 1868, and all breeds had registry associations by 1880. The purposes of these organizations are to preserve the purity of the breed through registration; to keep official records of the breed; to improve the breed through production testing programs, type classification, and dairy cattle shows; and to promote the breed.

The introduction of the dairy herd improvement associations (DHIA) in 1905 to measure the production of cows greatly aided the improvement of breeds, especially nonpurebred animals. These testing programs produced the needed data for the selection of replacement animals and for culling low-producing animals. The principal features of each of the five dairy breeds are given below.

Holstein-Friesian. This is the largest breed in number and in body size. A mature dairy cow weighs about 1500 lb (675 kg) and the bull 2200 (990 kg). The breed was developed in the North Holland and West Friesland provinces of the Netherlands. Holsteins

A champion milk producer, Harborcrest Rose Milly, the all-time, all-American Holstein-Friesian aged cow. (*Owned by Paclamar Farms, Louisville, Colorado*)

were first imported to the United States in 1621 and 1625 by the Dutch settlers; however, there are no descendants of these animals in the country. In the mid-1850s cattle were imported into Massachusetts from which descendants are now present. Importations in large numbers were made up to 1905.

The Holstein is a distinct black and white, is noted for its rugged conformation, and has a decided advantage in the production of beef and veal, as well as milk. The breed produces the largest volume of milk with the lowest milk fat percentage, averaging about 3.7%. Since the late 1940s the percentage of Holstein cattle in the United States has increased, because the milk-pricing systems have favored breeds that produce large volumes of milk with relatively low butterfat. One of the outstanding cows in the breed (see **illus**.) was classified Excellent with 97 points, had four records of over 23,000 lb (10,350 kg) of milk on twice-a-day milking, and had a lifetime production of over 138,000 lb (62,100 kg) of milk.

Jersey. The Jersey is the smallest of the dairy breeds. The mature cow weighs about 1000 lb (450 kg) and mature bull 1500 lb (675 kg). The Jerseys were developed on the Island of Jersey, one of the Channel Islands between France and England. The breed was first imported into the United States in 1850, with large importations occurring between 1870 and 1890. The percentage of Jerseys has been decreasing primarily due to the milk-pricing systems and a decreased consumption of butter. The Jersey ranges from a light fawn to black; some animals have white markings. The breed has the highest milk-fat content, averaging 4.9%, and a high solids-not-fat content. The milk is notably yellow, which is due to a high carotene content.

Guernsey. The Guernsey breed is slightly larger than the Jersey and has many of the latter's common characteristics. The mature female weighs 1100 lb (495 kg) and mature bull 1700 lb (765 kg). The Guernsey varies from an almost red to a very light fawn, with various sizes of light markings. The breed was developed on the Island of Guernsey, another Channel Island. They were first imported into the United States in 1830; while some importations still occur, they have decreased in numbers. The Guernseys average 4.7% milk fat and have a high solids-not-fat content. The milk also is a deep yellow due to the high carotene content.

Ayrshire. The Ayrshire breed was developed in the County of Ayr in Scotland around 1750. The mature cow weighs approximately 1200 lb (540 kg) and the bull 1850 lb (830 kg). The Ayrshire is red and white and noted for its general beauty, straightness of top line, symmetry of udder, ruggedness, and excellent grazing ability. The Ayrshires were first imported from Scotland to Massachusetts in 1837, with frequent importations made until 1860.

The average milk-fat content of the Ayrshire is 3.9%. The breed is known for its longevity and lifetime production rather than for its high lactation records.

Brown Swiss. The Brown Swiss is one of the larger breeds with the mature cow weighing about 1400 lb (630 kg) and the mature bull 2000 lb (900 kg). The Brown Swiss varies from a light fawn to almost black, with a mouse color predominating. Brown Swiss have a quiet temperament but a tendency to be stubborn. It is the slowest-maturing breed and is noted for its longevity. Because of size and ruggedness, they are well adapted for beef and veal purposes.

Brown Swiss is one of the oldest breeds. It was developed in the mountain territory of northeastern Switzerland, where in the winter the cows are housed in valleys and barns and in the spring are pastured on mountain slopes. This developed the ruggedness and excellent grazing ability. The cattle (less than 200) were first imported into the United States in 1869. The milk averages 4.1% milk fat and is well adapted to market milk, condensed milk, and cheese.

Dairy cattle feeding. Research on dairy cattle feeding has provided information on feeding animals according to their requirements to the dairy industry, interested in meeting nutritional needs at a minimum cost. The nutrient requirements of the dairy cow depend upon a daily maintenance need, milk production and milk-fat test, pregnancy, and growth. The two most important requirements to be met are energy and protein. Various energy standards are used, but the commonest ones in the United States are the total digestible nutrient (TDN) and net energy (NE) standards. The mineral requirements of the cow, particularly calcium, phosphorus, sodium, chloride, and iodine, are another concern of the dairy farmer. The other minerals and vitamins are found in sufficient quantities in the feeds.

The main components of the dairy ration are hay, silage, pasture, and grain. Grain is the most expensive portion of the ration and is fed according to the milk production of the cow. Hay and silage are usually fed to the cows free-choice, and a cow consumes daily about 2 lb of hay or 6 lb of silage per 100 lb of body weight. This provides enough energy and protein to meet the maintenance requirements of the cow plus 10–20 lb (5–10 kg) of milk each day. The rest of the energy and protein must come from grain. Grass and corn are used as silage. *See* ANIMAL FEEDS.

Dairy cattle breeding. The breeding program involves two major aspects. One is the selection and breeding of cows to be used in the dairy herd. The second is reproduction and maintenance of a regular calving interval for maximum milk production. Many commercial dairy producers place most emphasis on production in the breeding program and select for type only if the conformation of the animal presents management problems, such as a poorly attached udder. Some purebred dairy cattle breeders who sell replacement stock emphasize the overall type of animal.

To make progress in a breeding program, the dairy producer must use records. DHIA records are most commonly used for the selection of milk production. Selection for type is based on official classification scores conducted by the breed associations or the dairy farmer's own appraisal. Most selection in the breeding program is done at two times. One occurs in the selection of the bull or bulls to be used to breed cows for development of future herd replacements; the second is selecting animals to be culled and replaced by new animals. Approximately 32% of the cows leave the herd each year, because of disease, reproductive difficulties, and low production. With the advent of artificial insemination (AI) and the use of sire proving programs by AI associations, desirably proven bulls are available almost everywhere. Approximately 55% of the dairy cows in the United States are bred by artificial insemination. *See* BREEDING (ANIMAL).

Maintaining a good breeding program requires that cows calve at regular intervals. Most dairy workers try to have a 13–14-month calving interval. The ideal is 1 year; however, this is usually not possible. Cows calving at 13–14-month intervals produce more milk during their lifetimes than cows calving at wider intervals.

Dairy cattle management. Almost all cows in the United States are milked by machine. The recommended milking procedure involves a series of steps based upon the physiology of the cow, as well as the production of clean milk and disease control. The usual recommended steps in a milking operation are: Wash the cow's udder with a disinfectant solution and a paper towel, use a strip cup to detect abnormal milk, apply the milking machine 1–2 min after the cow's udder has been washed, remove the machine as soon as milk flow stops, and dip the teat ends in a disinfectant solution. Good milking procedures are required to maintain a low incidence of mastitis, which is the most costly dairy cattle disease. *See* AGRICULTURAL SCIENCE (ANIMAL); DAIRY MACHINERY; PASTEURIZATION.

G. H. Schmidt

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Dairy machinery

Equipment used in the production and processing of milk and milk products, including milking machines, cream separators, coolers, pasteurizers, homogenizers, butter-making equipment, evaporators and dryers, and related items of equipment. The equipment must be easy to clean and designed to prevent contamination of the milk or milk products from dirt, oil, soluble metals, insects, and other foreign materials.

Construction and design. Stainless steel, an alloy of chromium and steel, is widely used and is highly satisfactory for direct contact with milk and other food products. If properly used, stainless steel does not affect the flavor and is corrosion-resistant to food products. However, corrosion of stainless steel may be caused by prolonged contact with food or by removal of the protective oxide layer, which must be maintained to provide corrosion resistance. The layer is removed by prolonged contact with chlorine. Stainless steel surfaces must be cleaned regularly after use with detergent solutions, and should be sanitized before, rather than after, use so that excessive chlorine contact will be avoided. Surfaces are sanitized by using a hypochlorite solution of approximately 200 ppm of available chlorine. *See* STAINLESS STEEL.

Most modern dairy equipment is designed to be cleaned in place (CIP), without disassembly, by pumping detergents and cleaning solutions through the entire system. Equipment is specially designed to avoid pockets, to provide smooth surfaces, and to avoid burning-on or buildup of the product on food-contact surfaces.

Certain standards of design, construction, operation, and sanitation of milk and food equipment, known as the 3-A standards, have been established; they were developed by three associations and are widely used throughout the industry. The 3-A standards specify such items as composition of the material, thickness of material (gage), surface finish, design of corners and joints, size, power requirements, and method of utilization.

The science of hydraulics is basic to the fluid-flow processes of pumping, piping, agitating, mixing, centrifuging, and homogenizing. *See* HYDRAULICS.

Electric motors are used to supply power for these processes. Electricity is used for heating for small equipment and for instrumentation and control of processing equipment. Boilers furnish hot water and steam for cleaning (equipment and plant) and pasteurizing operations. Steam is the principal utility used for heating. Heating is a major utility necessary for pasteurization, sterilization, sanitation, and production of evaporated and dried milk. Refrigeration is provided through direct-expansion systems, by sweet water (cooled tap water) systems, or with brine systems for cooling of incoming product, storage, ice cream production, and for control of plant environment.

Refrigeration and heating require the major portion of the energy in dairy and food plants. In a medium-size plant the following basic utilities are

required per 1000 gal (3.78 kiloliters) of processed product: 130 kWh of electricity, 72 ton-h of refrigeration, 350 lb (158 kg) of steam, and 8000 lb (3630 kg) of water. *See* REFRIGERATION; THERMODYNAMIC PRINCIPLES.

Milking machines. These units extract milk from the cow's udder and deliver the milk to an adjacent container or, for a pipeline milking system, directly to a central cooling tank. Teat cups extract milk by intermittent vacuum action on a flexible inner wall called an inflation. A strip cup should be used before milking to check the condition of the milk. Machine milking requires 3.5–5 min, which is less time than for hand milking. A milking machine consists basically of a unit for providing a vacuum with a power unit, pipes or hoses to the animal, a device for providing pulsations, and a container for collecting the product.

A piston or vane-type pump, usually driven by an electric motor, is used for developing a vacuum of 10–15 in. Hg (34–51 kilopascals); a vacuum storage tank next to the outlet of the vacuum pump reduces the fluctuations of the piston. Intermittent action of 45–55 cycles/min is secured by the pulsator unit. Milk may be pumped or transferred to a central point by pipeline using a vacuum. *See* VACUUM PUMP.

Types. Milking machines can be classified as portable and pipeline units. The motor and vacuum pump of the portable unit can be either a part of the milker head or on a cart which is moved along behind the cows. The pipeline type can include either a pipeline for the air supply from the vacuum pump to the milker head or a pipeline from the cow to the milk house for transport of milk. With the latter type, there is usually a separate air line. About 1.5 kWh per cow per month is required for electricity for milking using a portable unit.

Operation. Two hoses go to the teat-cup units. One hose, called the air hose, extends from the pulsator, which alternately provides air at atmospheric pressure and vacuum (low pressure) to the teat cups. The other hose is the milk hose, which is under vacuum and moves the milk, caused to flow from the udder at atmospheric pressure, from the teat cups to the milker pail, which is under vacuum, or to the pipeline for a pipeline milking machine (**Fig. 1**). The rubber inflation surrounds the teat and provides the action for milking by the air-vacuum relationship set up by the pulsator. The rubber inflation is set in the metal cup, and the space between the inflation and the metal cup is alternately connected to air and vacuum, thus resulting in a massaging action. The space inside the inflation is always under vacuum when the teat cups are attached to the cow. The uniform vacuum on the inside of the inflation holds the teat cups in place.

Sanitization. Thorough cleansing of all equipment parts in contact with milk is essential to the production of milk of acceptably low bacteria count. Washing should be done immediately after using by first rinsing with cold water, then hot water using a wetting agent, rinsing with boiling water, and dismantling of equipment. A sanitizing solution may be run through the reassembled equipment just before milking. Local and federal public health ordinance requirements must be met for equipment and for the milk house.

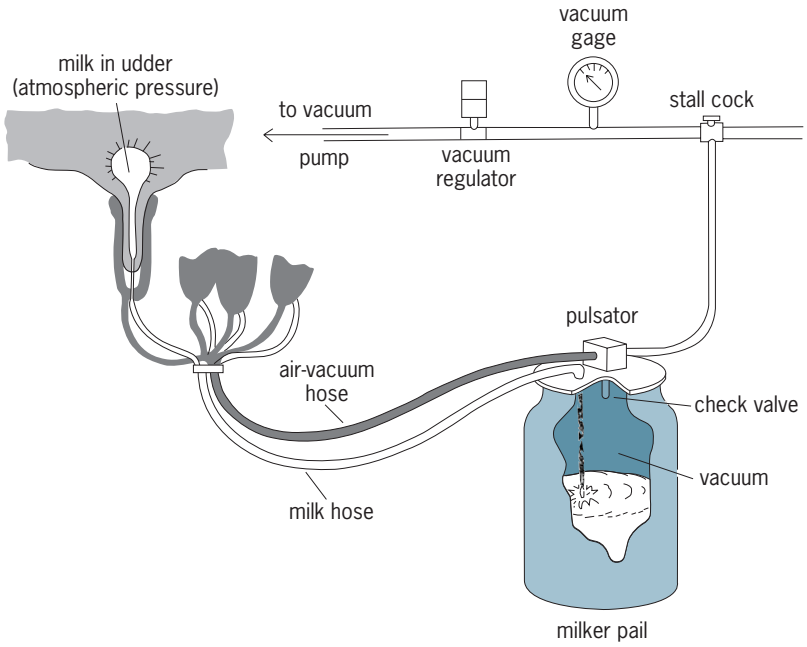

Fig. 1. Arrangement of the component parts of a milking machine.

Farm milk coolers. Milk must be rapidly cooled after production to avoid rapid growth of microorganisms in the product. Milk is normally about 93–95°F (34–35°C) as it leaves the cow. Most ordinances require that milk be cooled to 45–55°F (7–13°C) within 2 or 3 h after milking.

Surface coolers may be used for cooling the milk coming from the cow before it is put in the tanker. The surface cooler may be supplied with cool well water with the final cooling done by mechanical refrigeration. For milk which has been cooled over a surface cooler, the cans may be placed in refrigerated storage to maintain the temperature. Can coolers may be used for cooling the milk coming from the cow to below 50°F (10°C). Water is sprayed over the surface of the can or the can is immersed in cold water to provide the cooling. Bulk milk coolers are now widely used for refrigerating and holding milk on farms. Tanks meeting 3-A standards cool milk to 50°F (10°C) within 1 h and to 40°F (4.5°C) within 2 h after milking. The temperature of the product should not rise above 50°F when warm milk is added. The milk is normally held at 40°F or below. Milk is normally held 1 or 2 days in the farm bulk tanks (**Fig. 2**).

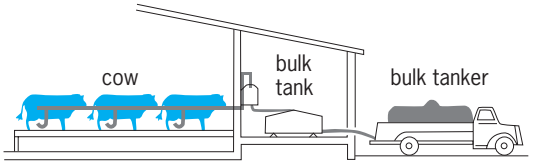

Fig. 2. Handling of milk from cow to tanker.

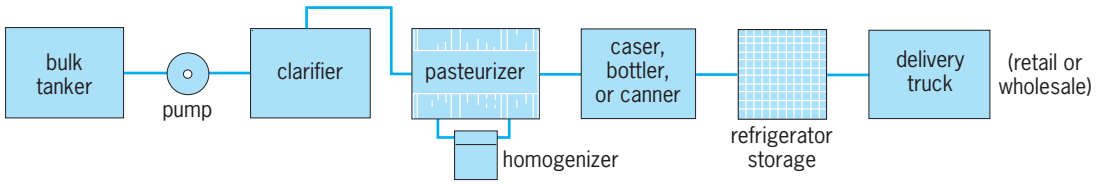

Fig. 3. Processing plant operations.

Bulk milk tankers. Tankers for bulk milk have insulated cylindrical tanks mounted in a horizontal position, as a trailer or on a truck, and are used for pickup of bulk milk. The tanker driver is responsible for sampling, weighing, checking quality and temperature, and handling the milk. Tolerances for measuring the quantity of product have been established by the National Bureau of Standards. The milk is pumped from the farm bulk tank to the tanker to be delivered to the processing plant. At the plant, a pump is usually used for emptying the tanker so that the milk can be transferred from the bulk tanker to a storage tank in the plant. The principal operations in the processing plant are shown in **Fig. 3**.

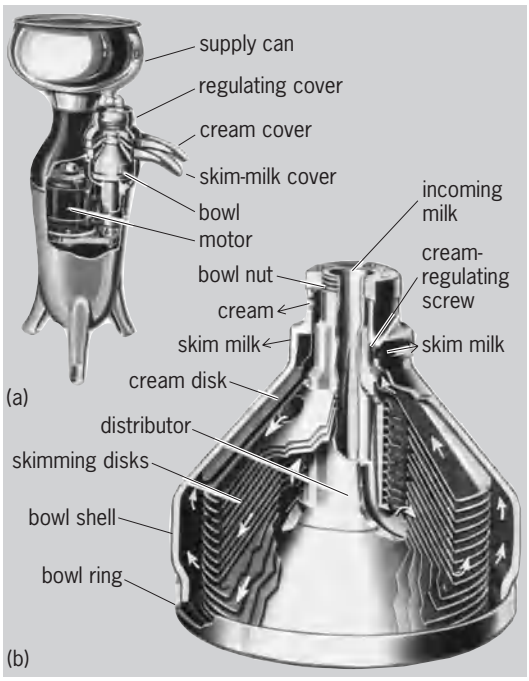

Fig. 4. Cream separator driven by electric motor. (*a*) Entire unit. (*b*) Cross-sectional view of the separating bowl. (*DeLaval Separator Co.*)

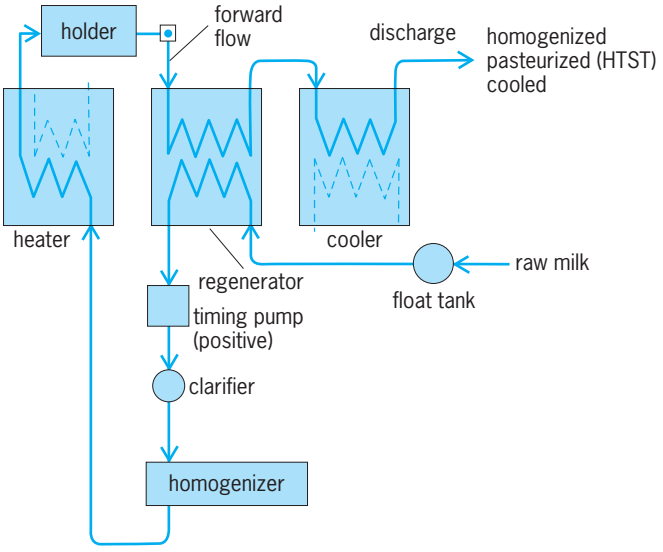

Fig. 5. Schematic representation of pasteurization processes. Details of bypass, relief lines, equalizer, and check valves are not included.

Clarifier. A clarifier removes foreign particles from milk by centrifugal force. A filter is sometimes used for this operation but is less efficient than a clarifier. In appearance and operation the clarifier is similar to a separator, but instead of a cream spout, which is on a separator for removal of a portion of the product, the clarifier accumulates the sediment in the bowl. The sediment must be removed periodically in order to maintain product flow.

The milk is normally clarified before homogenization in the bottled product so that the pasteurized product will not become contaminated. Clarification may be at temperatures of 50–95°F (10–35°C). Generally, a lower temperature of clarification is desired to avoid possible foaming from incorporation of air, to eliminate extraneous material before it is dissolved in the milk, thereby preventing off-flavors, and to reduce removal of components of the product and buildup of sediment in the centrifugal bowl.

Separator. Mechanical centrifugal devices used to extract fat from milk are called cream separators. Milk normally contains 3–6% fat in the form of globules which are lighter in density than the fluid skim milk. These particles rise to the surface when fresh milk is allowed to stand quiescent for several hours and form a layer of cream which may be skimmed from the surface.

In the centrifugal cream separator (**Fig. 4**), milk is moved into the bottom of an airtight bowl turning at 6000–10,000 rpm, which subjects the milk to radial centrifugal forces up to 500 *g*. The bowl contains stacked disks in the form of inverted cones which divide the milk into thin layers and provide the fluid friction necessary to bring the milk to the needed rotational speed. The angle of the inverted cone is such that the lighter cream particles tend to be forced up the center, where they are removed, and the heavier skim milk flows down the outside of the bowl and into the top, where it is collected. An efficient separator will leave no more than 0.1% fat in the milk. *See* CENTRIFUGATION.

The usual method of centrifugally separating the cream from the milk is to heat the milk to at least 90°F (32°C) and as high as 160°F (71°C). As the temperature of the milk increases, the viscosity decreases and

more efficient separation is obtained. Separators specially designed for efficient separation of cold milk have been developed.

Homogenizer. A homogenizer is a high-pressure pump, which is adjustable in pressure, with three or five pistons, and which forces the milk through small openings to break up and reduce the size of the fat globules. The size of the particles is regulated in general by adjusting the pressure of the valve. One- or two-stage units can be used, with the product passing through one or two homogenizing valves. Pressures from 2000 to 2500 lb/in.^{2} (14 to 17 megapascals) are normally used for single-valve milk homogenizers.

Milk must be heated before homogenization to inactivate lipase activity and to provide efficient breakup of the fat particles. The homogenizer may also be used as a pump.

Pasteurizer. This type of equipment is used to heat milk to a predetermined temperature and hold it at that temperature long enough to kill the pathogenic organisms that may be present, and, equally important, to inactivate enzymes. Satisfactory pasteurization can be obtained at various time-temperature combinations, ranging from at least 30 min at no less than 145°F (63°C), for batch pasteurization, to at least 15 s at no less than 161°F (72°C), for high-temperature short-time (HTST) systems. Flash pasteurization at 230°F (110°C), followed by discharge into a vacuum chamber for rapid cooling, is also used. This proce-dure is also termed the vacuum process. Batch-type pasteurizers are heated externally with hot water and require thorough agitation of the milk to heat the milk uniformly. Continuous-flow pasteurizers of the HTST type use regeneration as a means of increasing thermal efficiency by transferring heat from the hot outgoing milk to the cool incoming milk (**Fig. 5**). Cream and milk products to which sugars or other materials have been added must be given a higher temperature or longer time of pasteurization.

The trend is to higher temperatures and shorter times. Sterilized milk not needing refrigeration is now being produced. Sterilizing milk followed by aseptic packaging provides a means of minimum heat treatment and maximum product quality as contrasted with sterilization in the can.

Pasteurized products are placed in glass, metal, paper, or plastic cartons. Some products are placed in bulk containers which are usually placed in cases that are stacked automatically and moved by conveyor into storage. The storage is maintained at 40°F (4.5°C) or less. Milk products are often stored in the cooler, in truck lots, so that the retail or wholesale delivery trucks can rapidly make up their loads from the cooler storage. Automatic equipment for stacking and unstacking of cases, for emptying and filling cases with bottles or cartons or both, and for inspecting the cleanliness of containers is used.

Milk is submitted to a vacuum treatment to maintain a uniform flavor and odor by removal of low-boiling-point volatile components. Feed flavors and variation of product throughout the year, often caused by volatiles, are removed by submitting the product to a vacuum of 10–20 in. Hg (34–68 kPa) in one or two units. If used, this equipment is normally a part of the overall system of milk pasteurization.

Evaporation and drying. Water is removed from milk to reduce the bulk and decrease the cost of transportation and storage.

Evaporator. Evaporated milk is produced by removing water to condense the product to 3:1 or 2:1 volume ratios. Condensed milk is often treated by adding sugar. A vacuum pan is operated at a temperature of 130–140°F (55–60°C) or at 25 in. Hg (85 kPa) vacuum. For large operations, multiple-effect evaporators with one, two, or three stages of moisture removal are used. Decreased steam cost can be obtained with multiple-effect evaporators, a reduction that must be balanced against increased equipment cost. Evaporation usually precedes drying operations. The cost of removing water from the product is less in the evaporator than in the dryer.

Dryer. Dryers are generally of two types: the drum, or roller, dryer and the spray dryer (**Fig. 6**). The drum

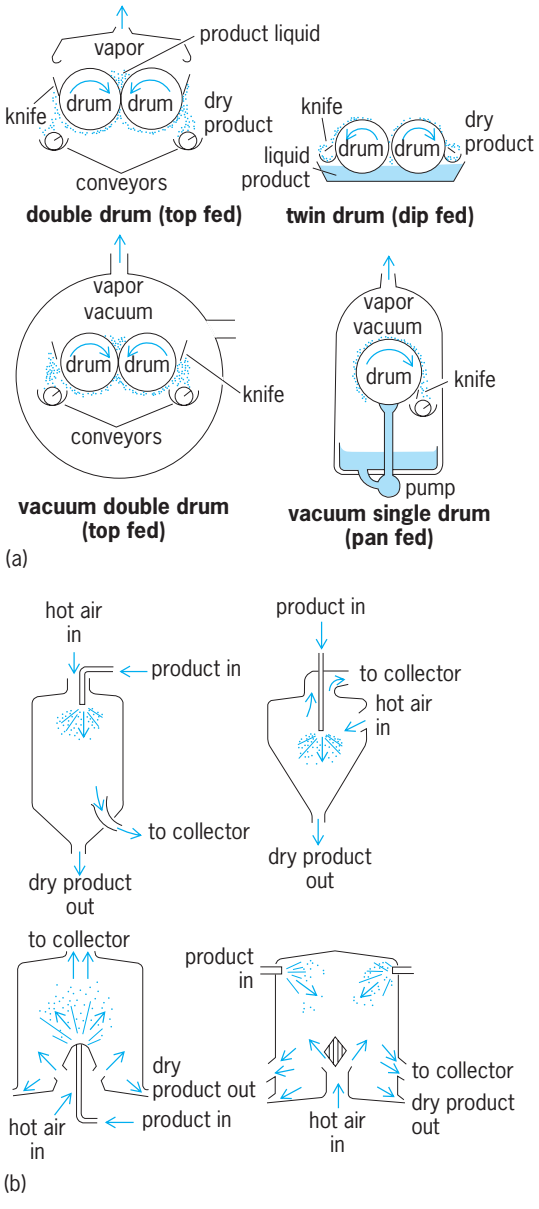

Fig. 6. Types of milk dryers and method of operation. (*a*) Drum dryers. (*b*) Vertical chamber spray dryers.

dryer is less expensive for moisture removal but has a greater heat effect on the product. With the drum dryer a thin film of product is moved over the steam-heated drum. As the drum turns, moisture is removed and a knife scrapes the product from the drum.

The spray dryer has an atomizing nozzle or a centrifugal device that breaks the product into small droplets. Heated air is forced around the atomizing head and evaporates moisture from the droplets. The dry product is then removed from the dryer. The spray-dried product has better solubility and provides, in general, a more satisfactory product for food. Instantizing equipment provides a procedure whereby dry particles are wetted slightly and agglomerates of particles formed. These agglomerates are then dried. The agglomerates have much more rapid solubility than the agglomerated products and provide many products that are known in the trade as instant products or instant powders. Improvements in methods of processing and drying milk have led to rapid increase in utilization of spray-dried and instantized products for the beverage milk industry. This practice will lead to wide distribution of milk products throughout the United States and the world. *See* BUTTER; CHEESE; DAIRY CATTLE PRODUCTION; ICE CREAM; MILK.

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D'Alembert's paradox

A theorem in fluid mechanics which states that no forces act on a body moving at constant velocity in a straight line through a large mass of incompressible, inviscid fluid which was initially at rest (or in uniform motion). This seemingly paradoxical theorem can be understood by first realizing that inviscid fluids do not exist. If such fluids did exist, there would be no internal physical mechanism for dissipating energy into heat; hence there could be no force acting on the body, because work would then be done on the fluid with no net increase of energy in the fluid.

The viscosity of many fluids is very small, but it is essential in explaining the forces that act on bodies moving in them. The action of viscosity creates a rotation of the fluid particles that come near the surface of a moving body. This vorticity, as it is called, is convected downstream from the body so that the assumption or irrotationality of the fluid motion, made in the proof of D'Alembert's theorem, does not correspond to reality for any known fluid. For a winglike body this viscous action sets up a circulation around the body which creates a lifting force. Viscosity also gives rise to tangential stresses at the body surface, called skin friction, which result in a drag force on the body. The work done on the fluid by moving a body through it shows up first as kinetic energy in the wake behind the body, which is gradually dissipated into heat by further action of viscosity.

D'Alembert's theorem does not preclude the possibility of a couple acting on the body, and in fact the irrotational, inviscid fluid theory does predict such a couple. This couple is almost always such as to cause the body to present its greatest projected area in the direction of motion. *See* FLUID-FLOW PRINCIPLES.

Arthur E. Bryson, Jr.

D'Alembert's principle

The principle that the resultant of the external forces $\mathbf{F}$ and the kinetic reaction acting on a body equals zero. The kinetic reaction is defined as the negative of the product of the mass m and the acceleration $\mathbf{a}$. The principle is therefore stated as $\mathbf{F} - m\mathbf{a} = 0$. While D'Alembert's principle is merely another way of writing Newton's second law, it has the advantage of changing a problem in kinetics into a problem in statics. The techniques used in solving statics problems may then provide relatively simple solutions to some problems in dynamics; D'Alembert's principle is especially useful in problems involving constraints. *See* CONSTRAINT.

If D'Alembert's principle is applied to the plane motion of a rigid body, the techniques of plane statics can be used. The principal advantage is that in a dynamics problem the torques must be calculated about a fixed point or about the center of mass, while in statics the torques can be calculated about any point.

Paul W. Schmidt

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Dalitz plot

Pictorial representation in high-energy nuclear physics for data on the distribution of certain three-particle configurations. Many elementary-particle decay processes and high-energy nuclear reactions lead to final states consisting of three particles (which may be denoted by a, b, c, with mass values m_a, m_b, m_c). Well-known examples are provided by the K-meson decay processes, Eqs. (1) and (2), and by the K- and $\overline{K}$-meson reactions with hydrogen, given in Eqs. (3) and (4). For definite total energy

$$K^+ \rightarrow \pi^+ + \pi^+ + \pi^- \tag{1}$$

$$K^+ \rightarrow \pi^0 + \mu^+ + \nu \tag{2}$$

$$K^+ + p \rightarrow K^0 + \pi^+ + p \tag{3}$$

$$K^- + p \rightarrow \Lambda + \pi^+ + \pi^- \tag{4}$$

E (measured in the barycentric frame), these final states have a continuous distribution of configurations, each specified by the way this energy E is shared among the three particles. (The barycentric frame is the reference frame in which the observer

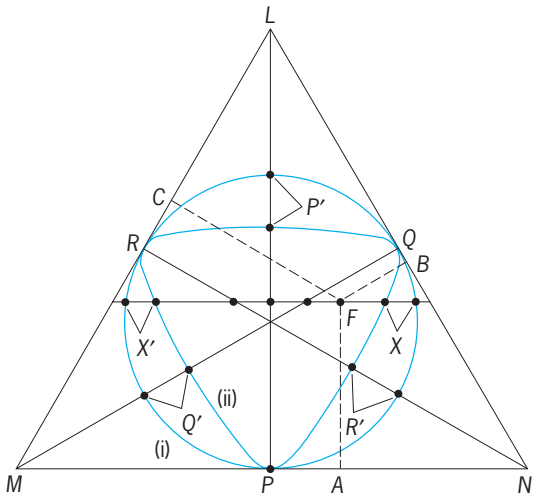

Fig. 1. A three-particle system (*abc*) in its barycentric frame is specified by a point *F* so that perpendiculars *FA*, *FB*, *and FC* to the sides of an equilateral triangle *LMN* (of height *Q*) are equal in magnitude to the kinetic energies T_a, T_b, T_c, where *Q* denotes their sum. See Eq. (5). Curve (i) encloses all points *F* which correspond to physically allowed configurations, for equal masses and nonrelativistic kinematics; curve (ii) corresponds to curve (i) when relativistic kinematics are used, appropriate to the decay process $\omega(785\ \text{MeV}) \rightarrow \pi^+\pi^-\pi^0$.

finds zero for the vector sum of the momenta of all the particles of the system considered.) *See* ELEMENTARY PARTICLE.

Equal mass representation. If the three particles have kinetic energies T_a, T_b, and T_c (in the barycentric frame), Eq. (5) is obtained. As shown in **Fig. 1**,

$$T_a + T_b + T_c = E - m_a c^2 - m_b c^2 - m_c c^2 = Q \quad (5)$$

this energy sharing may be represented uniquely by a point F within an equilateral triangle LMN of side $2Q/\sqrt{3}$, such that the perpendiculars FA, FB, and FC to its sides are equal in magnitude to the kinetic energies T_a, T_b, and T_c. This exploits the property of the equilateral triangle that ($FA + FB + FC$) has the same value (Q, the height of the triangle) for all points F within it. The most important property of this representation is that the area occupied within this triangle by any set of configurations is directly proportional to their volume in phase space. In other words, a plot of empirical data on this diagram gives at once a picture of the dependence of the square of the matrix element for this process on the a, b, c energies.

Not all points F within the triangle LMN correspond to configurations realizable physically, since the a, b, c energies must be consistent with zero total momentum for the three-particle system. With nonrelativistic kinematics (that is, $T_a = p_a^2/2m_a$, etc.) and with equal masses m for a, b, c, the only allowed configurations are those corresponding to points F lying within the the circle inscribed within the triangle, shown as (i) in Fig. 1. With unequal masses the allowed domain becomes an inscribed ellipse, touching the side MN such that $MP{:}PN$ equals $m_b{:}m_c$ (and cyclical for NL and LM). More generally, with relativistic kinematics ($T_a = \sqrt{(m_a^2C^4 + p_a^2c^2} - m_ac^2$, etc.), the limiting boundary is distorted from a simple ellipse or circle. This is illustrated in Fig. 1 by the boundary curve (ii), drawn for the $\omega \rightarrow 3\pi$ decay process, where the final masses are equal. This curve was also calculated by E. Fabri for Eq. (1), and this plot is sometimes referred to as the Dalitz-Fabri plot. In the high-energy limit $E \rightarrow \infty$, where the final particle masses may be neglected, the boundary curve approaches a triangle inscribed in LMN.

The following points (and the regions near them) are of particular interest:

1. All points on the boundary curve. These correspond to collinear configurations, where a, b, c have parallel momenta.
2. The three points of contact with the triangle LMN. For example, point P corresponds to the situation where particle c is at rest (and therefore carries zero orbital angular momentum).
3. The three points which are each farthest from the corresponding side of the triangle LMN. For example, point P' on Fig. 1 corresponds to the situation where b and c have the same velocity (hence zero relative momentum, and zero orbital angular momentum in the bc rest frame).

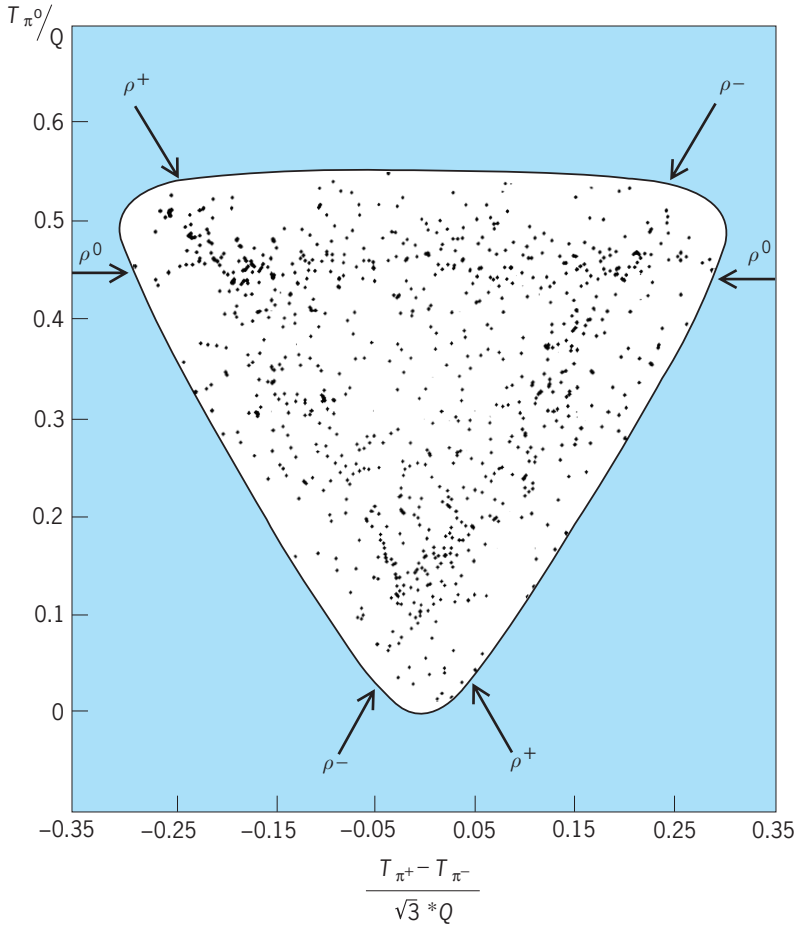

Fig. 2. Dalitz plot for 823 examples of the antiproton annihilation process, $p + p \rightarrow \pi^+ + \pi^+ + \pi^0$. Arrows show positions expected for ρ(765)-meson resonance bands, which appear clearly for the ($\pi^+\pi^-$), ($\pi^-\pi^0$), and ($\pi^0\pi^+$) systems. Distribution is symmetrical between the six sectors obtained by drawing the three axes of symmetry. The authors interpret this distribution as being due to (1) the reaction $\bar{p}p \rightarrow \rho\pi$ occurring in the $I = 0$ $S_1{}^3$ initial state, (2) the reaction $pp \rightarrow 3\pi$ (*s*-wave pions) occurring in the $I = 1$ $S_0{}^1$ initial state, with roughly equal intensities. (*After C. Baltay et al., Annihilation of antiprotons in hydrogen at rest, Phys. Rev., 140:B1039, 1965*)

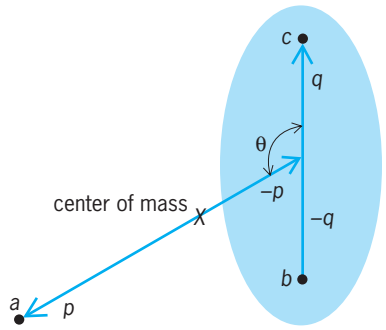

Fig. 3. Coordinate system for relativistic three-particle system. For given total energy *E*, the two momenta, *q* and *p*, are related in magnitude by the following equations: $E = \sqrt{(m_a^2 + p^2)} + \sqrt{(M_{bc}^2 + p^2)}$ and $M_{bc} = \sqrt{(m_b^2 + q^2)} + \sqrt{(m_c^2 + q^2)}$.

If the process occurs strongly through an intermediate resonance state, say $a + (bc)^*$ where $(bc)^* \rightarrow b + c$, there will be observed a "resonance band" of events for which T_a has the value appropriate to this intermediate two-body system. Such a resonance band runs parallel to the appropriate side of the triangle [the side *MN* for the case $(bc)^*$, and cyclically] and has a breadth related with the lifetime width for the resonance state.

Antiproton annihilation. The Dalitz plot shown for equal masses in Fig. 1 has been especially useful for three-pion systems, since it treats the three particles on precisely the same footing. Points placed symmetrically with respect to the symmetry axis *PL* represent configurations related by the interchange of π_b and π_c. The symmetry axes *PL*, *QM*, and *RN* divide the allowed regions into six sectors; the configurations in each sector can be obtained from those corresponding to one chosen sector (for example, the sector such that $T_a \geq T_b \geq T_c$) by the six operations of the permutation group on three objects. These operations are of particular interest for three-pion systems, since pions obey Bose statistics; the intensities in the six sectors are related with the permutation symmetry of the orbital motion in the three-pion final state. The Dalitz plot shown in **Fig. 2**, for the antiproton capture reaction $\bar{p}p \rightarrow \pi^+\pi^-\pi^0$, illustrates these points. Three ρ-meson bands [$\rho(765\ \text{MeV}) \rightarrow \pi\pi$] are seen, corresponding to intermediate states $\pi^+\rho^-$, $\pi^0\rho^0$, and $\pi^-\rho^0$; the six sectors have equal intensity. *See* BOSE-EINSTEIN STATISTICS.

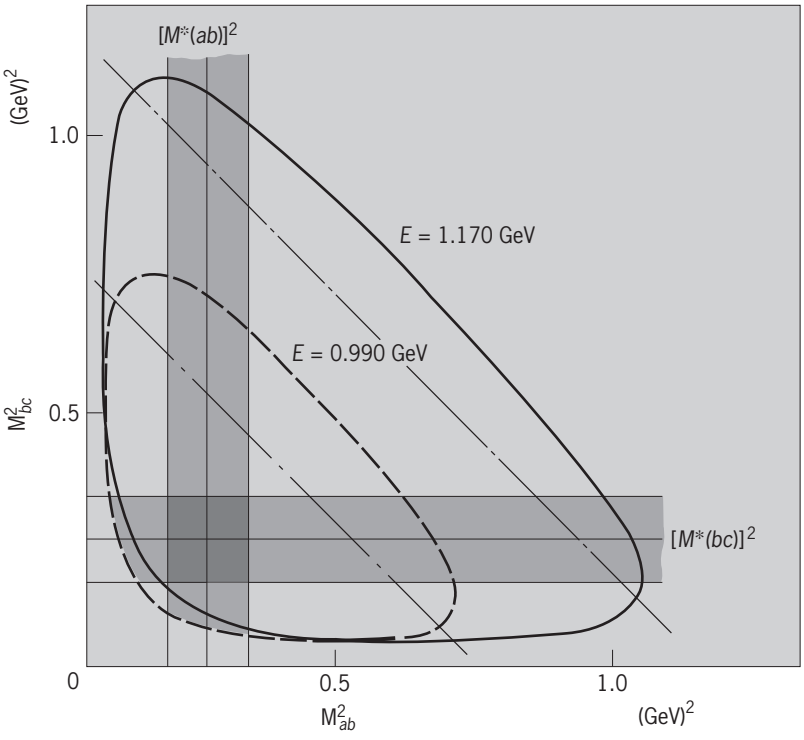

Fig. 4. An unsymmetrical Dalitz plot. Configuration of system (*abc*) is specified by a point (M^2_{ab}, M^2_{bc}) in a rectangular coordinate system, where M_{ij} denotes the barycentric energy of the two-particle system (*ij*). Kinematic boundaries have been drawn for equal masses $m_a = m_b = m_c = 0.14$ GeV and for two values of total energy *E*, appropriate to a three-pion system ($\pi^+\pi^-\pi^+$). Resonance bands are drawn for states (*ab*) and (*bc*) corresponding to a (fictitious) π-π resonance mass 0.5 GeV and full width 0.2 GeV. The dot-dash lines show the locations a (*ca*) resonance band would have for this mass of 0.5 GeV, for the two values of the total energy *E*.

Relativistic three-particle system. **Figure 3** depicts a less symmetric specification for a three-particle system. The momentum of *c* in the (*bc*) barycentric frame is denoted by **q**, the momentum of *a* in the (*abc*) barycentric frame by **p**, and the angle between **p** and **q** by θ. For fixed energy T_a, the points *F* on Fig. 1 lie on a line parallel to *MN*; as cos θ varies from +1 to −1, the point *F* representing the configuration moves uniformly from the left boundary *X*′ to the right boundary *X*. If cartesian coordinates are used for *F*, with origin *P* and *y* axis along *NM* (as has frequently been found useful in the literature), then Eqs. (6) hold. Note that, for fixed *x*, *y* is linearly related with cos θ.

$$x = T_a/Q \qquad y = (T_b - T_c)/Q\sqrt{3} \tag{6}$$

Unsymmetrical plot. The Dalitz plot most commonly used is a distorted plot in which each configuration is specified by a point with coordinates (M^2_{ab}, M^2_{bc}) with respect to right-angled axes. This depends on the relationship given in Eq. (7) and its

$$M_{bc}^2 = (E - m_a c^2)^2 - 2ET_a \tag{7}$$

cyclic permutations, for the total barycentric energy M_{bc} of the two-particle system *bc*. This plot may be obtained from Fig. 1 by shearing it to the left until *LM* is perpendicular to *MN*, and then contracting it by the factor $\sqrt{3/2}$ parallel to *MN* [which leads to a cartesian coordinate system (T_c, T_a), finally reversing the direction of the axes [required by the minus sign in Eq. (7)] and moving the origin to the point $M^2_{ab} = M^2_{bc} = 0$. The plots shown in **Fig. 4** correspond in this way to the relativistic curve (ii) in Fig. 1 for two values of the total energy *E*. This distorted plot retains the property that phase-space volume is directly proportional to the area on the plot. As shown in Fig. 4, the $(ab)^*$ and $(bc)^*$ resonance bands have a fixed location on this plot; data from experiments at different energies *E* can then be combined on the same plot to give a stronger test concerning the existence of some intermediate resonance state. On the other hand, the $(ca)^*$ resonance bands run across the plot at 135° and move as *E* varies, so that a different

choice of axes [say (M_{ca}^2, m_{ab}^2)] is more suitable for their presentation.

Conclusion. It must be emphasized that the Dalitz plot is concerned only with the internal variables for the system (*abc*). In general, especially for reaction processes such as Eqs. (3) and (4), there are other variables such as the Euler angles which describe the orientation of the plane of (*abc*) relative to the initial spin direction or the incident momentum, which may carry additional physical information about the mechanism for formation of the system (*abc*). The Dalitz plots usually presented average over all these other variables (because of the limited statistics available); this sometimes leads to a clearer picture in that there are then no interference terms between states with different values for the total spin-parity. However, it is quite possible to consider Dalitz plots for fixed values of these external variables, or for definite domains for them. *See* EULER ANGLES; GOLDHABER TRIANGLE. Richard H. Dalitz

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Dallis grass

A general term for a genus of grasses of which the most important species is the deeply rooted perennial *Paspalum dilatatum*. Dallis grass is widely used in the southern United States, mostly for pasture, and remains productive indefinitely if well managed. Dallis grass does best on fertile soils and responds to lime and fertilizer. On heavier soils it remains green throughout the winter unless checked by heavy frosts. Seed production is hampered by infection with ergot, a fungus that invades developing seeds and produces purplish-black horny bodies. Ergot-bearing seed heads are very toxic to livestock, whether in pasture or in hay, and Dallis grass must be so managed as to prevent consumption of infected heads by livestock. *See* CYPERALES; ERGOT AND ERGOTISM; FERTILIZER; GRASS CROPS; PLANT PATHOLOGY. Howard B. Sprague

Dalton's law

The total pressure of a mixture of gases is the sum of the partial pressures of each gas in the mixture. The law was established by John Dalton (1766–1844). In his original formulation, the partial pressure of a gas is the pressure of the gas if it alone occupied the container at the same temperature. Dalton's law may be expressed as $P = P_A + P_B + \cdots$, where P_J is the partial pressure of the gas J, and P is the total pressure of the mixture; this formulation is strictly valid only for mixtures of ideal gases. For real gases, the total pressure is not the sum of the partial pressures (except in the limit of zero pressure) because of interactions between the molecules.

In modern physical chemistry the partial pressure is defined as $P_J = x_J P$, where x_J is the mole fraction of the gas J, the ratio of its amount in moles to the total number of moles of gas molecules present in the mixture. With this definition, the total pressure of a mixture of any kind of gases is the sum of their partial pressures. However, only for an ideal gas is the partial pressure (as defined here) the pressure that the gas would exert if it alone occupied the container. *See* AVOGADRO'S NUMBER; GAS; KINETIC THEORY OF MATTER; THERMODYNAMIC PRINCIPLES. P. W. Atkins

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Dam

A barrier or structure across a stream, river, or waterway for the purpose of confining and controlling the flow of water. Dams vary in size from small earth embankments for farm use to high, massive concrete structures for water supply, hydropower, irrigation, navigation, recreation, sedimentation control, and flood control. As such, dams are cornerstones in the water resources development of river basins. Dams are now built to serve several purposes and are therefore known as multipurpose (**Fig. 1**). The construction of a large dam requires the relocation of existing highways, railroads, and utilities from the river valley to elevations above the reservoir. The two principal types of dams are embankment and concrete. Appurtenant structures of dams include spillways, outlet works, and control facilities; they may also include structures related to hydropower and other project purposes. *See* ELECTRIC POWER GENERATION; IRRIGATION (AGRICULTURE); WATER SUPPLY ENGINEERING.

Dams have been built since ancient times, and improvements were made at varying intervals as engineering technology developed. However, very rapid advances occurred in the twentieth century as a result of developments in the use of concrete, soil mechanics, and construction equipment. In the early 1900s, concrete dams became thinner, and a new era of thin arch dams began. Earth and rock-fill embankment dams became economical during and after World War II. In 1980, an innovative method of using earth-moving and compacting equipment to place dry concrete (roller-compacted concrete) greatly improved the economics of concrete dams. As of December 1997, there was a total of 100,781 dams in the United States. Most of these are small recreational projects and farm ponds. Of this total, 6389 dams have a height of 50 ft or more and 1586 dams have a height of 100 feet or more. Numerous dams have been constructed in various countries worldwide (**Table 1**). Many dams possess considerable height, volume, and reservoir capacity (**Table 2**).

Purposes. Dams are built for specific purposes. In ancient times, they were built only for water supply or irrigation. Early in the development of the United States, rivers were a primary means of transportation, and therefore navigation dams with locks (**Fig. 2**) were constructed on the major rivers. Dams

Fig. 1. John Day Lock and Dam, looking upstream across the Columbia River at Washington shore. In the foreground the navigation lock is seen, then the spillway beyond it, and then the powerhouse. The John Day multiple-purpose project has the highest single-lift navigation lock in the United States. (*U.S. Army Corps of Engineers*)

have become more complex to meet large power demands and other needs of modern countries (**Figs. 3** and **4**). Although recreation is a popular purpose of small private dams, it is planned as a benefit with an assigned monetary value in federal projects in the United States. A typical summary of purposes of nonfederal dams is shown in **Table 3**.

Features. In addition to the standard impounded reservoir and the appurtenant structures of a dam (spillway, outlet works, and control facility), a dam with hydropower requires a powerhouse, penstocks, generators, and switchyard. The inflow of water into the reservoir must be monitored continuously, and the outflow must be controlled to obtain maximum benefits. Under normal operating conditions, the reservoir is controlled by the outlet works, consisting of a large tunnel or conduit at stream level with control gates. Under flood conditions, the reservoir is maintained by both the spillway and outlet works.

The reservoir level of a flood control dam is maintained as low as possible to create the maximum amount of storage space for use in the flood season. For an irrigation project, the reservoir is filled as high as possible in the winter and early spring, and it is maintained at that level for maximum release of

Fig. 2. Lock and Dam 1 on the Mississippi River between St. Paul and Minneapolis, Minnesota. Built in 1917 with a single lock to provide barge traffic to Minneapolis, the dam failed in 1929 and was rebuilt with twin locks in 1932. (*U.S. Army Corps of Engineers*)

TABLE 1. Countries with the largest number of dams*

Country	Number of dams†
China	18,595
United States	6,389
Japan	2,142
India	1,085
Spain	690
Republic of Korea	628
Canada	580
United Kingdom	529
Brazil	489
Mexico	487
France	432
Italy	408
Australia	374

*From "ICOLD Register of Dams," updated 1996.
†Dams at least 50 ft (15 m) high.

TABLE 2. Landmark dams of the world

Order	Name	River	Country	Type*	Height ft	Height m	Year completed†
			HIGHEST DAMS				
1	Rogun	Vakhsh	Tajikistan	E-R	1099	335	1985
2	Nurek	Vakhsh	Tajikistan	E	984	300	1980
3	Xiaowan	Lancang	China	A	958	292	UC
4	Grand Dixence	Dixence	Switzerland	G	935	285	1961
5	Inguri	Inguri	Georgia	A	892	272	1980
6	Vajont	Vajont	Italy	A	860	262	1961
7	Manuel M. Torres	Grijalva	Mexico	R	856	261	1980
8	Tehri	Bhagirathi	India	E-R	856	261	2006
9	Alvaro Obregon	Tennasco	Mexico	G	854	260	1946
10	Mauvoisin	Drance de Bangnes	Switzerland	A	820	250	1957
11	Sayano-Shushensk	Yenisei	Russia	A	804	245	1989
12	Alberto Lieraso	Guavio	Colombia	R	797	243	1990
13	Mica	Columbia	Canada	E	794	242	1972
14	Ertan	Yangtze/Yalong	China	A	787	240	1998
15	La Esmeralda	Bata	Colombia	R	777	237	1975
16	Kishau	Tons	India	G	774	236	UC
17	Oroville	Feather	United States	E	770	335	1968
18	El Cajon	Comayagua	Honduras	A	768	234	1984
19	Chirkei	Sulak	Russia	A	764	233	1978
20	Shuibuya	Qingjiang	China	R	764	233	UC

			LARGEST EMBANKMENT DAMS BY VOLUME		Volume 10^3 yd^3	Volume 10^3 m^3	Year completed
1	Chapeton	Paraná	Argentina	E-G	387,400	296,200	UC
2	Pati	Paraná	Argentina	E-G	311,411	238,100	UC
3	Tarbella	Indus	Pakistan	E-R	194,223	148,500	1967
4	Fort Peck	Missouri	United States	E	125,624	96,050	1937
5	Lower Usuma	Usuma	Nigeria	E	121,635	93,000	1990
6	Cipasang	Cimanuk	Indonesia	E-R	117,842	90,100	UC
7	Tucurui	Tocantis	Brazil	E-R	84,098	85,200	1984
8	Ataturk	Firat	Turkey	R	111,172	85,000	1990
9	Yacyreta	Parana	Paraguay/Argentina	E	105,944	81,000	1997
10	Guri	Caroni	Venezuela	R-G	102,014	78,000	1986
11	Rogun	Vakhsh	Tajikistan	E-R	98,746	75,500	1985
12	Oahe	Missouri	United States	E	91,996	70,339	1958
13	Parambikulam	Parambikulam	India	E-G	90,461	69,165	1967
14	Mangla	Jhelum	Pakistan	E	85,497	65,370	1967
15	Gardiner	S. Saskatchewan	Canada	E	85,014	65,000	1968
16	Afsluitdijk	Zuiderzee	Netherlands	E	82,921	63,400	1932
17	Mangla	Jhelum	Pakistan	E	82,893	63,379	1967
18	Oroville	Feather	United States	E	77,997	59,635	1968
19	San Luis	San Luis	United States	E	77,897	59,559	1967
20	Nurek	Vakhsh	Tajikistan	E	75,858	58,000	1980

			LARGEST ARTIFICIAL RESERVOIRS IN TERMS OF CAPACITY		Capacity 10^3 acre ft‡	Capacity 10^5 m^3	Year completed
1	Owen Falls	Victoria Nile	Uganda		166,033	296,200	1954
2	Kariba	Zambezi	Zimbabwe/Zambia		146,090	180,200	1959
3	Bratsk	Angara	Russia		137,010	169,000	1964
4	High Aswan	Nile	Egypt		131,335	162,000	1970
5	Akosombo	Volta	Ghana		119,952	147,960	1965
6	Daniel Johnson	Manicouagan	Canada		115,000	141,851	1968
7	Guri	Caroni	Venezuela		109,446	135,000	1986
8	Krasnoyarsk	Yenisei	Russia		59,425	73,300	1967
9	WAC Bennett	Peace	Canada		57,000	70,309	1967
10	Zeya	Zeya	Russia		55,452	68,400	1978
11	Cabora Bassa	Zambezi	Mozambique		51,075	63,000	1974
12	La Grande 2	La Grande	Canada		50,033	61,715	1978
13	Chapeton	Paraná	Argentina		49,129	60,600	UC
14	La Grande 3	La Grande	Canada		48,659	60,020	1981
15	Ust-Ilim	Angara	Russia		48,075	59,300	1977
16	Boguchany	Angara	Russia		47,183	58,200	UC
17	Kuibyshev	Volga	Russia		47,021	58,000	1955
18	Serra da Mesa	Tocantins	Brazil		44,103	54,400	UC
19	Caniapiscau Barrage KA 3	Caniapiscau	Canada		43,608	53,790	1980
20	Upper Wainganga	Wainganga	India		41,103	50,700	UC

*E = embankment, earth-fill. R = embankment, rock-fill. E-R = embankment, earth and rock-fill. G = gravity. A = arch.
†UC = under construction.
‡1 acre foot = 1 acre of water 1 foot deep.

Fig. 3. Guri Dam, Venezuela, showing construction under way on the second stage of the concrete gravity dam. Embankment dams are shown on the left and right abutments. The concrete batch plant is shown at left with a trestle for transporting concrete shown on the downstream face of the dam. The second powerhouse with an excavated outlet channel is shown at the center of the photo. (*C.V.G. Electrification del Caroni*, *C. A. EDELCA*, *Caracas*, *Venezuela*)

water during the dry season. The reservoir level of a hydropower dam is maintained as constant as feasible to create a uniform head for power generation. Water quality is an important ingredient in sustaining a balance in nature and is taken into account in modern dam design, construction, and operation. The chemical quality and temperature of the water are monitored in the reservoir. Intake ports at various depths allow selective withdrawal and mixing to produce the desired temperature and oxygen content, in order to enhance downstream environmental conditions. Fish ladders, that is, stepped series of elevated pools, are provided at many dams to allow free passage of fish upstream and downstream. Screens are used to keep fish out of the turbines. *See* RESERVOIR.

The discharge from modern dams must be managed carefully and continuously. During floods, reservoir inflows may exceed maximum discharges and cause reservoir levels to rise. To prevent dams from overtopping and possibly failing, spillways are provided to pass floodwater safely. They are commonly built at elevations just below dam crests and without gates. These uncontrolled, ungated spillways are designed to allow all of the excess water to pass.

Fig. 4. Itaipu Dam, on the Paraná River between Brazil and Paraguay. An operating spillway with a flip bucket on the Paraguay side of the river is shown at left. The hollow concrete gravity dam with powerhouse construction at the downstream toe is shown at center. (*G. S. Sarkaria*, *International Engineering Company*, *Inc.*)

TABLE 3. Primary purpose of nonfederal dams in the United States*

Purpose	Percent of total
Recreation	35
Flood control	15
Water supply	10
Irrigation	10
Farm and livestock water	15
Power	3

*From "National Dam Inventory," 1996 update.

Fig. 5. Aerial view of North Fork Dam, a combination earth-and-rock embankment of the North Fork of Pound River, Virginia. The channel-type spillway (left center) has a simple overflow weir. (*U.S. Army Corps of Engineers*)

In other cases, spillways are constructed at even lower levels and contain gates that are operated from the control facilities. The tops of these gates are lower than the dam crests, thus allowing some control of floodwater.

Penstocks (usually steel pipes or concrete-lined tunnels) are used to convey water from the reservoir through or around the dam to the powerhouse. The penstocks are connected to the turbines, and the water flow is controlled by valves. The number and size of the penstocks vary, depending on the number of generators and amount of water needed.

All the features of a dam are monitored and operated from a control room. The room contains the necessary monitors, controls, computers, emergency equipment, and communications systems to allow project personnel to operate the dam safely under all conditions. Standby generators and backup communications equipment are necessary to operate the gates and other reservoir controls in case of power failure. Weather conditions, inflow, reservoir level, discharge, and downstream river levels are also monitored. In addition, the control room monitors instrumentation located in the dam and appurtenant features that measures their structural behavior and physical condition.

Requirements. All dams are designed and constructed to meet specific requirements. First, a dam should be built from locally available materials when possible. Second, the dam must remain stable under all conditions, during construction, and ultimately in operation, both at the normal reservoir operating level and under all flood and drought conditions. Third, the dam and foundation must be sufficiently watertight to control seepage and maintain the desired reservoir level. Finally, it must have sufficient spillway and outlet works capacity as well as freeboard to prevent floodwater from overtopping it.

Types. Dams are classified by the type of material from which they are constructed. In early times, the materials were earth, large stones, and timber, but as technology developed, other materials and construction procedures were used. Most modern dams fall into two categories: embankment and concrete. Embankment dams are earth or rock-fill; other gravity dams and arch and buttress dams are concrete.

Earth-fill dam. Earth is the predominant material in this type of embankment dam. Earth dams are further classified by the construction method: hydraulic-fill or rolled-fill. A hydraulic-fill dam is one in which the soil is excavated, transported, and placed by flowing water. A large dredge operating in the river or other borrow area pumps a slurry of earth and water to the damsite. Here the coarse-grained materials settle on the outside portion of the embankment, and the remaining slurry is allowed to pond at the center, where the very fine-grained clay-size particles settle to form the impervious portion of the dam.

Advances in earth-moving construction equipment during World War II led to widespread construction of rolled-earth-fill dams. Economic advantages of this type of embankment often include the use of material available from the site excavation as embankment material, and the ready availability of fill material at or near the damsite. Other advantages of earth-fill dams include their adaptability to a wide variety of site configurations and their tolerance of weak foundations.

At various stages during excavation and placement of the fill, the moisture content of the soil may be adjusted by wetting or drying in order to optimize its performance in the finished embankment. The soil is spread on the embankment in uniform layers 8–12 in. (20–30 cm) thick and compacted with sheepsfoot or rubber-tired rollers. The rollers make from four to eight passes, depending on the desired

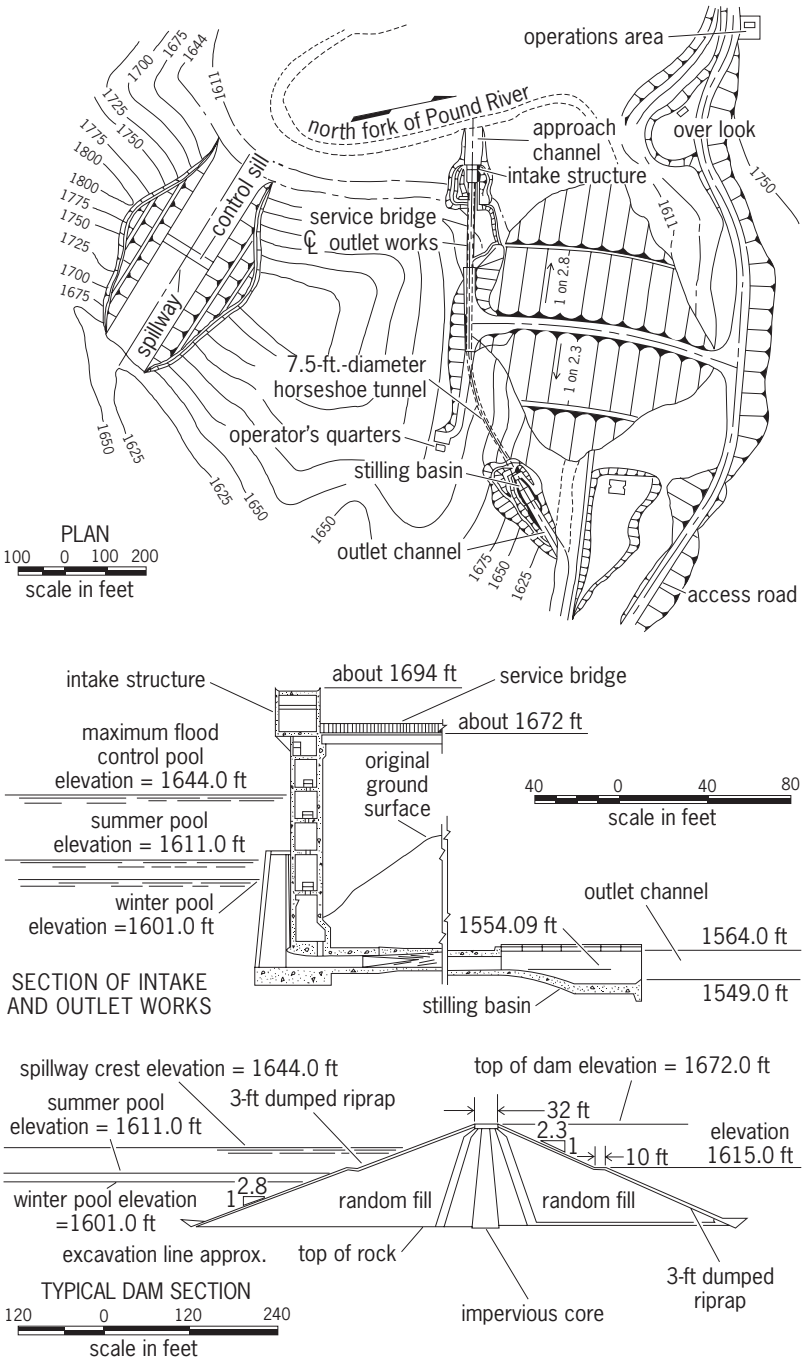

Fig. 6. Plan and sections of North Fork of Pound Dam, Virginia. 1 ft = 0.3 m. (*U.S. Army Corps of Engineers*)

density. Typical dry densities for rolled earth fill range 100–130 lb/ft^3 (1602–2083 kg/m^3).

Seepage control is an important aspect of earth-dam design. In early times, earth embankments were homogeneous, and seepage would emerge on the downstream slope just above ground level. If uncontrolled, such seepage can move soil particles and cause failure. In 1940, filter criteria were developed after careful scientific tests on all types of soil and on the sands and gravel to be used as filter material. These criteria allow engineers to design internal drains to collect and remove seepage. When an earth dam is built on a site where bedrock is at considerable depth, the foundation must be treated to control seepage. Typical treatment includes one or a combination of several things: upstream impervious blanket, cutoff wall, drainage blanket, gravel drains excavated into the foundation at the downstream toe, and relief wells. *See* FOUNDATIONS.

Earth-fill dams are by far the most popular type in the world. They make up 78%, or 27,260, of all those dams at least 50 ft (15 m) high. The earth dam's spillway is usually located in adjacent terrain rather than in the dam itself. The outlet works are either a conduit in the valley or a tunnel in one of the abutments (**Figs. 5** and **6**). Excavation for the spillway and outlet works usually produces large quantities of rock. As a result, the use of both earth and rock in an embankment is a common practice.

Rock-fill dam. A rock-fill dam is a rolled fill embankment composed of fragmented rock with an impervious zone or membrane located on the upstream face or incorporated in the center of the embankment. The impervious membrane is typically a concrete slab or asphalt layer on the upstream face. The impervious zone is typically a thin internal core of earth fill. The earth core is separated from the rock shell (the structural mass of the dam) by zones of small rock fragments or gravel, to prevent the earth from washing into the rock fill, and a drain to control seepage. In Europe, it is common to use asphalt for the impervious zone.

Rock-fill dams require solid rock foundations and sites where large quantities of rock are available. Seepage through rock foundations is prevented or minimized by grout curtains. Rock-fill dams are usually more economical than concrete gravity dams at sites having wide valleys and adequate foundations. Spillways and outlet works are at locations similar to those of earth-fill dams. The primary advantage of rock-fill dams is that they require less material. Rock fill has a higher shear strength than earth fill and therefore permits steeper exterior slopes.

Rock for the embankment is normally excavated by drilling and blasting. Hole spacing and powder charges are set to produce a particular gradation of rock fragments for the dam. The rock is placed and spread in the same manner as earth but in thicker layers, 18–36 in. (45–90 cm). The material is normally compacted by weighted or vibratory steel drum rollers. Dry densities of rock fill are normally in the range 110–145 lb/ft^3 (1762–2323 kg/m^3).

Rock-fill dams became popular in the United States during the California gold rush in the 1860s and 1870s, when many dams were built in remote locations to store water for use in hydraulic sluicing. Of the 34,780 dams in the world that are at least 50 ft (15 m) high, 1590 are rock-fill embankments.

Concrete gravity dam. Concrete gravity dams are massive structures, characterized by vertical or near-vertical upstream faces and steep downstream faces (**Figs. 7** and **8**). They are designed with enough weight to resist being overturned or moved by the force of the water in the reservoirs. They are economical only at sites with shallow, high-strength rock foundations. Because of the large volumes of

concrete involved, adequate sources of high-quality aggregates must be available near the sites. Concrete is composed of water, cement, pozzolan, aggregates, and entrained air. These ingredients are proportioned to produce concrete of the desired workability, durability, and strength as economically as possible. The density of concrete in dams typically ranges 140–160 lb/ft^3 (2243–2563 kg/m^3). An important feature of gravity dams is the simplicity with which safe spillways and outlet works can be provided. *See* CONCRETE.

The design of a concrete gravity dam is controlled by stability considerations and internal stresses. The structure must be able to resist water, sediment, and ice pressures from the reservoir, as well as earthquake forces. Computers permit rapid solutions of complex equations for determining the magnitude and distribution of internal stresses. These dams are built in monolithic units by using the block method of construction (**Fig. 9**). This promotes dissipation of heat produced by hydration of the cement (chemical combination with the water) and thus helps minimize the volume changes associated with overheating that cause tensile stresses and cracking. The blocks are separated by construction joints. In building a block, the concrete is placed in horizontal layers and vibrated to eliminate voids. The monoliths are cast on top of firm rock foundations that have been cleaned with water and treated by placement of cement and water slurry or grout in the cracks and joints.

A concrete gravity dam usually contains an internal gallery large enough to allow for physical inspection and for collection of drainage from downstream drain holes drilled into the foundation. Grout holes to reduce seepage in the foundation are also drilled from the gallery in the vertical or upstream direction. The grout is injected under pressure to force it into all joints and openings encountered at depth.

Arch dam. The arch dam is a thin concrete dam that curves upstream from each abutment (**Figs. 10** and **11**). Such dams are classified as thin, medium, or thick arch, depending on the ratio of structural height to base thickness. The ratio is 0.2 or less for a thin arch, 0.25 for a medium arch, and 0.3 or greater for a thick arch. The arch transmits the water pressure and other loads directly to the abutments and foundation. It contains significantly less concrete than a concrete gravity dam of the same height and length. Relatively narrow canyons favor the use of arch dams.

The shape of early arch dams was controlled by construction materials available at the time, and by less sophisticated understanding of structural behavior and the way that loads were transmitted through the curved structures to the foundations. As a result, arches were simple masonry structures with curved alignments and near-vertical upstream faces. This type was popular among water companies supplying domestic and irrigation water.

Beginning in the 1900s, improved structural analysis and actual performance records led to the use of variable-thickness arch dams. Varying the thickness can reduce the volume of concrete required. Measurement of the physical properties of concrete began in the late 1920s. This led to improved design procedures and measurement of actual performance with such instruments as strain gages. The concept of working stresses emerged in the late 1920s. The double-curvature shape (curved top to bottom as well as transversely) emerged in the mid-1950s. Vertical curving and shaping of the arch improves stress distribution. Making the compressive stresses levels throughout the dam as close as possible to the maximum allowable stress results in the minimum volume of concrete. A symmetrical profile is desirable. This may require excavation on one abutment if the canyon is not symmetrical. The economic upper limit of the length-to-height ratio of an arch dam lies between 4:1 and 6:1.

Fig. 7. Green Peter Dam, a concrete gravity type on the Middle Santian River, Willamette River Basin, Oregon. A gate-controlled overflow-type spillway is constructed through the crest of the dam; the powerhouse is at the downstream toe of the dam. (*U.S. Army Corps of Engineers*)

Buttress dam. The buttress dam consists of two principal structural elements: a sloping upstream deck that retains the water, and buttress walls that support the deck and transmit the loads to the foundation. Traditionally, buttress dams have been classified into three categories: flat slab, multiple arch, and massive head. The flat-slab type consists of a reinforced concrete flat slab inclined at about 45° and connected to buttresses. The multiple-arch type is a series of concrete arches spanning the buttresses. The massive-head type has a large mass of concrete in the section upstream from the buttresses.

In 1918, the flat-slab design was patented in the United States. About 200 buttress dams of all three categories have been built in the United States. Many are less than 150 ft (45 m) high. Some landmark

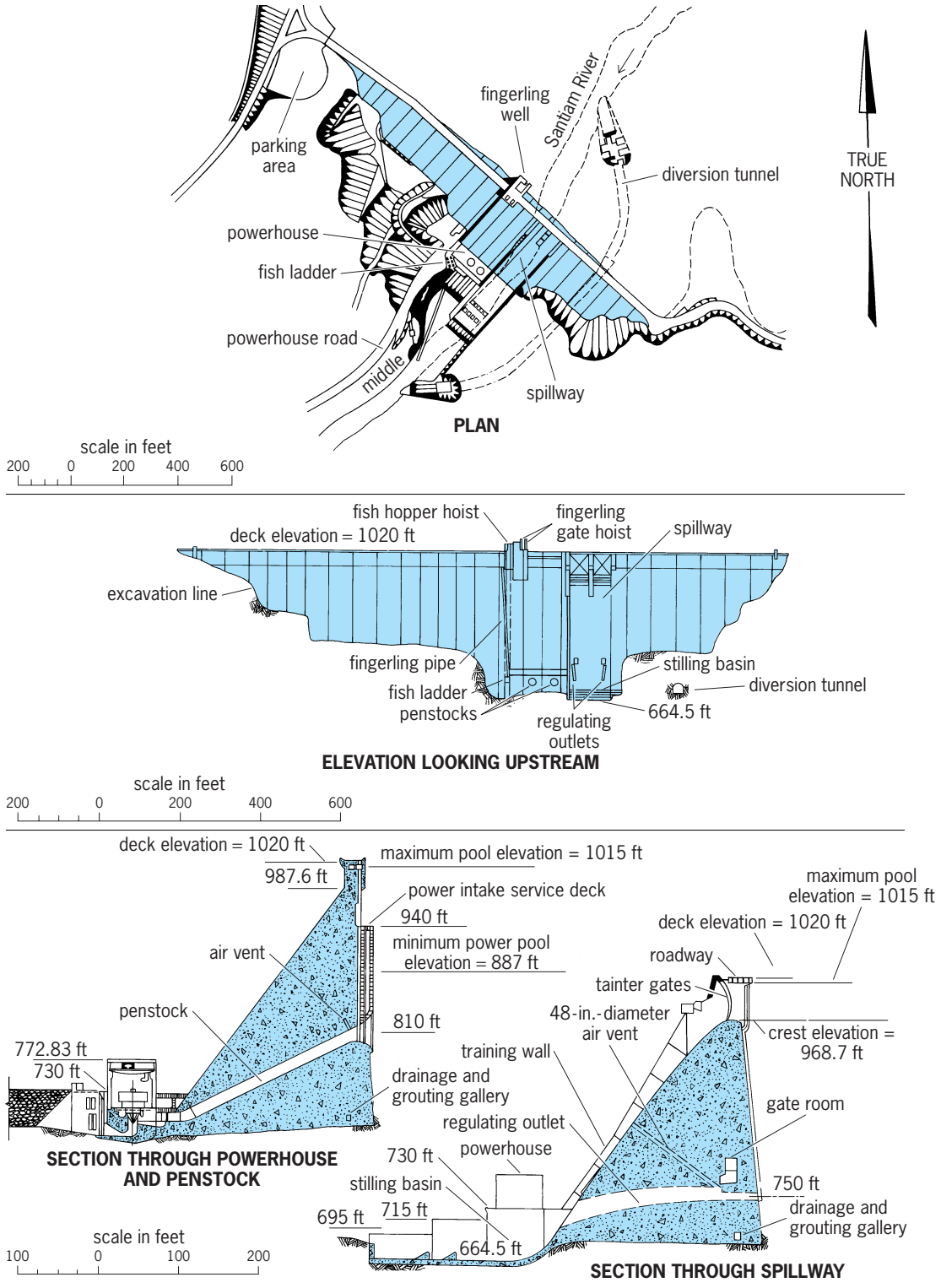

Fig. 8. Plan and sections of Green Peter Dam. 1 ft = 0.3 m, 1 in. = 2.5 cm. (*U.S. Army Corps of Engineers*)

buttress dams are the Daniel Johnson in Canada (1986), 702 ft (214 m) high, the world's highest multiple-arch buttress dam; and the José M. Oriol in Spain (1969), 426 ft (130 m), the world's highest flat-slab buttress dam.

Site and type selection. The type of dam for a particular site is selected on the basis of technical and economic data and environmental considerations. In the early stages of design, several sites and types are considered. Drill holes and test pits at each site provide soil and rock samples for testing physical properties. In some cases, field pumping tests are performed to evaluate seepage potential. Preliminary designs and cost estimates are prepared and reviewed by hydrologic, hydraulic, geotechnical, and structural engineers, as well as geologists. Environmental quality of

Fig. 9. Block method of construction on a typical concrete gravity dam. (*U.S. Army Corps of Engineers*)

the water, ecological systems, and cultural data are also considered in the site-selection process.

Factors that affect the type are topography, geology, foundation conditions, hydrology, earthquakes, and availability of construction materials. The foundation of the dam should be as sound and free of faults as possible. Narrow valleys with shallow sound rock favor concrete dams. Wide valleys with varying rock depths and conditions favor embankment dams. Earth dams are the most common type. *See* ENGINEERING GEOLOGY; FAULT AND FAULT STRUCTURES.

Construction process. Hydraulic-fill operations over a 4-year period at Fort Peck Dam, on the Missouri River, the largest embankment by volume in the United States (**Fig. 12**), dredged 156,000,000 yd^3 (119,340,000 m^3) of material. Of this volume, 122,000,000 yd^3 (93,333,000 m^3) was retained in the embankment. Large conventional excavation operations can produce hourly volumes of 2000–3000 yd^3 (1530–2293 m^3). Processing, hauling, placement, and compaction operations for earth or rock result in daily placement rates that vary from as low as 2500 yd^3 (1911 m^3) on small dams to 6500 yd^3 (4969 m^3) on larger dams.

The materials and construction procedures for concrete dams evolved gradually from the early dams in Asia and Europe to the modern massive concrete dams. Prior to 1900, portland cement used in the United States was imported from England. Thus, the early concrete dams built in the United States were masonry. Generally, the concrete was mixed and transported in wheelbarrows. In the case of cyclopean masonry, large irregular blocks of rock with mortar, small derricks were erected, and the maximum rate of placement approached a few hundred cubic yards a day. There was no attempt to cure the concrete. Between 1900 and 1930, concrete was placed by towers and chutes. Portland cement had become available in the United States, and placement rates improved. However, little attention was given to the mix design, and wet mixes that could easily flow in chutes were widely used.

Hoover Dam, on the Colorado River, was a major turning point in both the design and construction of concrete dams. Its unprecedented size,

Fig. 10. East Canyon Dam, a thin-arch concrete structure on the East Canyon River, Utah. There is an uncontrolled overflow-type spillway through the crest of the dam at the right. (*U.S. Bureau of Reclamation*)

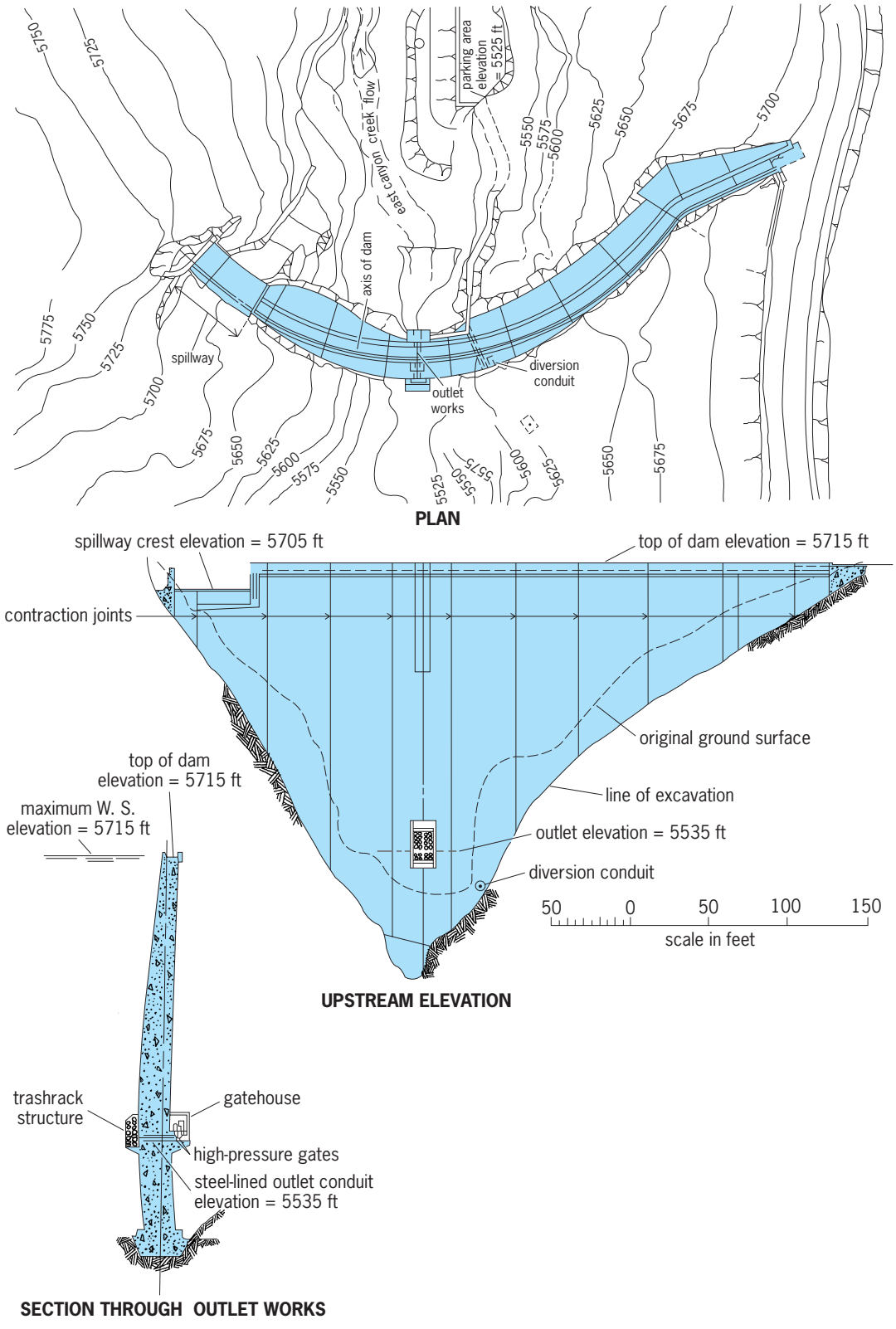

Fig. 11. Plan and sections of East Canyon Dam, Utah. 1 ft = 0.3 m. (*U.S. Army Corps of Engineers*)

4,400,000 yd^3 (3,363,800 m^3), led to the introduction of mass concrete placement. Average placement rates of 10,000 yd^3 (7645 m^3) per day were achieved. Advances in design resulting from the Hoover project led to the construction of Grand Coulee, on the Columbia River, 10,585,000 yd^3 (8,099,000 m^3). Two large concrete plants were used that supported a maximum placement of 20,680 yd^3 (15,810 m^3) per day and an average rate over the construction period of 6000 yd^3 (4587 m^3) per day.

Since 1980, the technology of placing dry concrete with paving equipment and compacting it with rollers has gained wide acceptance (**Fig. 13**). This construction method is known as roller-compacted concrete. By 1997, 30 dams at least 50 ft (15 m) high had been constructed in the United States using the roller-compacted concrete method. As of January 1998, there were 16 RCC dams in the world having a height of 100 m or greater. This method produces the high placement rates usually associated with earth-fill construction and results in economical structures. It was initially used in 1975 in the tunnel repairs at Tarbela Dam, Pakistan, and placement reached a maximum rate of 24,000 yd^3 (18,343 m^3).

An outstanding example of the rapid construction achieved by using roller-compacted concrete is Copperfield Dam in Australia. This 131-ft-high (40-m) dam contains 183,000 yd^3 (140,000 m^3) and required only 10 months from initial design to completion. The project was originally designed as an earth- and rock-fill dam, but it was switched to roller-compacted concrete for greater economy.

River diversion during construction. The designers of a dam must consider the stream flow around or through the damsite during construction. Stream flow records provide the information for use in determining the largest flood to divert during the selected construction period. One common practice for diversion involves constructing the permanent outlet works, which may be a conduit or a tunnel in the abutment, along with portions of the dam adjacent to the abutments, in the first construction period. In some cases, a temporary channel is built at a preferred diversion location, and levees are built to control the flow of water through the damsite. After the outlet works and main dam are completed to an appropriate level, the stream is diverted into the outlet works by a cofferdam high enough to prevent overtopping during construction. A downstream cofferdam is also required to keep the damsite dry. In the final construction period, the entire dam is brought to full height. *See* COFFERDAM.

Operation and maintenance. Personnel responsible for operation and maintenance of the dam become involved during the final design and construction to become familiar with design details that relate to operation. The operating instructions and maintenance schedule are published in a formal document for each dam. A schedule is established for collection and reporting of data for climatic conditions, rainfall, snow cover, stream flows, and water quality of the reservoir, as well as the downstream reaches. All these data are evaluated for use in reservoir regulation. Another schedule is established for the collection of instrumentation data used to determine the structural behavior and physical condition of the dam. These data are evaluated frequently.

Routine maintenance and inspection of the dam and appurtenant structures are ongoing processes. The scheduled maintenance is important to preserve the integrity of the mechanical equipment.

Periodic inspection and evaluation. Upon completion of construction, the project is inspected in detail by a team made up of the designers, construction managers, operations personnel, and other experts. The purpose is to ensure that the dam has been built as designed and can safely impound water and that all systems are ready for the initial reservoir filling and operation. In addition, the same team conducts an in-depth inspection once a year for about 5 years after completion and at 5-year intervals thereafter. Design criteria and performance of the dam as measured by instruments are reviewed during the life of the dam, and structural reanalyses are made when necessary. Photographs are taken to record rates of deterioration.

Fig. 12. Fort Peck Dam, Missouri River, Montana, the largest embankment dam by volume in the United States, 125,628,000 yd^3 (95,625,000 m^3). A hydraulic-fill dam, it was built between 1935 and 1939 for flood control, hydropower, irrigation, and navigation. (*U.S. Army Corps of Engineers*)

Fig. 13. Galesville Dam, Oregon, showing roller-compacted concrete construction; completed in 1985, it has a height of 157 ft (48 m) and a volume of 161,000 yd^3 (123,100 m^3).

The intake structure, trash racks, emergency gates, outlet conduit or tunnel, and stilling basin are normally under water and therefore require special procedures such as dewatering prior to the inspection. At normal velocity, the flowing water can severely erode soil and rock in the approach and discharge channels. High-velocity flow over small irregularities can cause a phenomenon known as cavitation, which can lead to rapid erosion of metal and concrete and can threaten the safety of the outlet works. *See* CAVITATION.

Instrumentation. As the technology of dam design and construction progressed, the need to measure performance and structural behavior became important in order to verify the design. Advances in instruments starting in the 1950s gave the designer a valuable tool. Instrumentation gave the engineer knowledge of how the temperatures from hydration in concrete varied and the effect on strength. Pressure cells were developed that gave information about the interaction between soil backfill and a concrete wall or structure as well as the actual load distribution. Piezometers (devices to measure water level), settlement plates, and slope indicators are used in measuring the performance of embankment dams. Plumb lines, strain gages, and uplift cells are used for the same purpose in concrete dams. In addition, instruments are used to measure vertical and horizontal movement, alignment and plumb, stresses, strains, water pressure, seismic effects, and the quantity and clarity of seepage.

Instrumentation for a dam is installed at first in the design phase to establish baseline data, then during construction and throughout the life of the dam as conditions warrant. The frequency with which instrumentation data are obtained is an extremely important issue and depends on operating conditions. Timely collection and evaluation of data are critical for periods when the loading changes, such as during floods and after earthquakes. Advances in applications of remote sensing to instrumentation have made real-time data collection possible. This is a significant improvement for making dam safety evaluations.

Safety. Throughout history there have been instances of dam failure and discharge of stored water, sometimes causing considerable loss of life and great damage to property. Failures have generally involved dams that were designed and constructed to engineering standards acceptable at the time. Most failures have occurred with new dams, within the first five years of operation.

As dam technology advanced with increasing knowledge of design principles and better understanding of foundation and material properties, dams became safer. There is no question that they can be built and operated safely. The major issue is to monitor deterioration as the structures and equipment get older. In earlier times, the sizes of spillways and outlet pipes had to be determined by judgment. As a result, overtopping was the main cause of dam failure. Little was known about soil mechanics and slope stability, and so slides and slope failures were common. Beginning in the 1930s, statistical methods were used to predict floods. Advances in soil mechanics in the later 1930s and early 1940s produced new methods of stability analysis that revolutionized the slope design for excavations and earth embankment dams. Historical data indicate that the causes of failure (in the order of their significance) are piping, overtopping, slope instability, conduit leakage (outlet works), and foundation failures. *See* PIPELINE; ROCK MECHANICS; SOIL MECHANICS.

TABLE 4. Some major dam failures in history

Dam	Country	Year
San Ildefonso	Bolivia	1626
Puentes	Spain	1802
Bradfield	England	1864
Johnstown	United States	1889
Walnut Grove	United States	1890
Bouzey	France	1895
Austin	United States	1911
Bila Densa	Czechoslovakia	1916
Tigra	India	1917
Glenco	Italy	1923
St. Francis	United States	1928
Alla S. Zerbimo	Italy	1935
Fred Burr	United States	1948
Malpasset	France	1959
Kuala Lumpur	Malaya	1961
Babi Yar	Soviet Union	1961
Baldwin Hills	United States	1963
Vaiont	Italy	1963
Buffalo Creek	United States	1972
Teton	United States	1976
Kelly Barnes	United States	1977
Machhu II	India	1979

It is estimated that about 150,000 dams around the world present a potential hazard to life or property; there have been 200 failures since 1900. Many of these have involved small dams. **Table 4** lists some major failures that resulted in considerable loss of life.

Dam failures cause loss of life and property damage in downstream reaches that are beyond the control of the dam owner or local government. For this reason, and because dam safety practices should apply to all dams, national governments have become involved in order to provide supervision and standardize regulations. The United States government published "Federal Guidelines for Dam Safety" in June 1979. This initiated a coordinated effort in management practices among federal agencies and set an example for private organizations that own dams. The International Commission on Large Dams (ICOLD) was formed in 1928 by 6 countries with the purpose of developing and exchanging dam design experience, and it has grown to 76 member countries. In 1982 ICOLD established a committee on dam safety to define common safety principles, integrate efforts, and develop guidelines, and in 1987 ICOLD published "Dam Safety Guidelines." Arthur H. Walz, Jr.

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Damping

A term broadly used to denote either the dissipation of energy in, and the consequent decay of, oscillations of all types or the extent of the dissipation and decay. The energy losses arise from frictional (or analogous) forces which are unavoidable in any system or from the radiation of energy to space or to other systems. For sufficiently small oscillations, the analogous forces are proportional to the velocity of the vibrating member and oppositely directed thereto; the ratio of force to velocity is $-R$, the mechanical resistance. For the role of damping in the case of forced oscillations, where it is decisive for the frequency response, *see* FORCED OSCILLATION; HARMONIC MOTION; MECHANICAL VIBRATION; OSCILLATION; RESONANCE (ACOUSTICS AND MECHANICS).

Damped oscillations. An undamped system of mass m and stiffness s oscillates at an angular frequency $\omega_0 = (s/m)^{1/2}$. The effect of a mechanical resistance R is twofold: It produces a change in the frequency of oscillations and it causes the oscillations to decay with time. If u is one of the oscillating quantities (displacement, velocity, acceleration) of amplitude A, then Eq. (1) holds in the damped case, whereas in the undamped case Eq. (2) holds. The reciprocal

$$u = Ae^{-\alpha t}\cos\omega_d t \tag{1}$$

$$u = A\cos\omega_0 t \tag{2}$$

time α in Eq. (1) may be called the damping constant.

In Eqs. (1) and (2), the origin for the time t is chosen so that $t = 0$ when $u = A$. The damped angular frequency ω_d in Eq. (1) is always less than ω_0; its value will be given later. According to Eq. (1), the amplitude of the oscillation decays exponentially; the time, given in Eq. (3), is that required for the amplitude to

$$1/\alpha = 2m/R \tag{3}$$

decrease to the fraction $1/e$ its initial value.

A common measure of the damping is the logarithmic decrement δ, defined as the natural logarithm of the ratio of two successive maxima of the decaying sinusoid. If T is the period of the oscillation, then Eq. (4) holds, so that Eq. (1) becomes Eq. (5). Thus

$$\delta = \alpha T \tag{4}$$

$$u = Ae^{-\delta t/T}\cos\omega_d t \tag{5}$$

$1/\delta$ is the number of cycles required for the amplitude to decrease by the factor $1/e$ in the same way that $1/\alpha$ is the time required.

The Q of a system is a measure of damping usually defined from energy considerations. In the present case, the stored energy is partly kinetic and partly potential; when the displacement is a maximum, the velocity is zero and the stored energy is wholly potential, while at zero displacement, the energy is wholly kinetic. The Q is π times the ratio of peak energy stored to energy dissipated per cycle. In the present example, this reduces to Eq. (6). The damped frequency ω_d (1) is related to the undamped frequency ω_0 of Eq. (2) by Eq. (7), so that for high-Q (lightly damped) systems, it is only slightly less than ω_0. *See* ENERGY.

$$Q = \omega_0 m/R = \pi/\delta \tag{6}$$

$$(\omega_d/\omega_0)^2 = 1 - (1/2Q^2) \tag{7}$$

Overdamping; critical damping. If α in Eq. (1) exceeds ω_0, then the system is not oscillatory and is said to be overdamped. If the mass is displaced, it returns to its equilibrium position without overshoot, and the return is slower as the ratio α/ω_0 increases. If $\alpha = \omega_0$ (that is, $Q = {}^1/_2$), the oscillator is critically damped. In this case, the motion is again nonoscillatory, but the return to equilibrium is faster than for any overdamped case.

Distributed systems. An undamped, one-dimensional wave of frequency $\omega/2\pi$ propagated in the positive direction of x is represented by Eq. (8), c being the velocity of the wave. If the vibration is maintained at $x = 0$ at the value $u = A\cos\omega t$, then the damping manifests itself as an exponential decrease of amplitude with distance x. Equation (8) is replaced by Eq. (9). The attenuation α' may depend on frequency. If the medium is terminated, the wave will be reflected from the ends, and a system of standing waves will be set up. Examples are a rod carrying sound waves, a piece of electrical transmission line or waveguide, and a vibrating violin string. Such a system has a number of natural frequencies $\omega_n/2\pi$, at each of which it behaves like the lumped system of the previous sections. The decay of a vibration is characterized by Q in Eq. (10), where $\lambda_n = 2\pi c/\omega_n$ is the wavelength.

$$u = A\cos\omega(t - x/c) \tag{8}$$

$$u = Ae^{-\alpha' x}\cos\left[\omega(t - x/c)\right] \tag{9}$$

$$Q = \omega_n/2\alpha' c = \pi/\alpha'\lambda_n \tag{10}$$

Hysteresis damping. At a given instant, the elongation (strain) of a metal bar which is under periodic, alternating stress is not determined exactly by the instantaneous value of the stress existing at that time. For example, the elongation is less at a given stress value when the stress is increasing than when it is decreasing. This phenomenon, which is known as

mechanical hysteresis, causes an undesirable energy loss. A vibration problem of serious nature exists in the blades of jet engines and other steam and gas turbines. The blade material itself exhibits a mechanical hysteresis damping which holds the vibrations in check. When the stress is small, the hysteresis damping is very small in all metals, but it rises suddenly when the stress reaches a certain value. Unfortunately, in most metals the stress at which hysteresis damping becomes large and that at which the metal fails because of fatigue are very close together. However, a much higher hysteresis damping at safe stresses than that of ordinary steel is exhibited by certain alloys. *See* HYSTERESIS; MECHANICAL VIBRATION; STRESS AND STRAIN.

Oscillating electrical circuits. A simple series electrical circuit consisting of an inductance L, resistance R, and capacitance C is exactly analogous to the mechanical system described by Eqs. (1) through (6). The inductance, resistance, and elastance $(1/C)$ correspond to the mass, mechanical resistance, and stiffness, respectively. A distributed electrical circuit, such as a section of transmission line or wave guide, is analogous to a vibrating rod or disk. *See* CIRCUIT (ELECTRICITY).

In the ordinary electrical oscillator, the frequency is controlled by an electrical resonator (tank) lumped at the lower, and distributed at the higher, frequencies. Good frequency stability is associated with a high-Q tank. For frequencies of tens of megahertz (MHz) and below, mechanical resonators can be constructed which have a much higher Q than the equivalent electrical tanks. Thus, very stable electrical oscillators have mechanical resonators as their frequency-determining elements. Such an electromechanical system can operate only if there is some coupling between the electrical and mechanical aspects of the system. *See* OSCILLATOR.

The coupling can be arranged in various ways. In some materials, such as quartz, the constitutive relations involve the mechanical and electrical variables jointly; thus, for example, an electric field may produce a strain in the absence of any stress. Thus, the coupling is inherent in the quartz itself. Quartz crystals are used for frequency control of oscillators in the range from kilohertz to perhaps 100 MHz; the Q of a high-frequency crystal may be several million. Some low-frequency oscillators are controlled by tuning forks having a Q of several hundred thousand, the action being similar to that of the electric bell or buzzer. *See* PIEZOELECTRICITY.

The electrostatic motor-generator effect provides the coupling in such mixed systems as the condenser microphone and electrostatic loudspeaker, and the electromagnetic motor-generator effect plays the same role in the dynamic microphone, dynamic loudspeaker, and in various electrical instruments. *See* ALTERNATING-CURRENT CIRCUIT THEORY; ELECTRIC TRANSIENT; ELECTROSTATICS; LOUDSPEAKER; MICROPHONE; RESONANCE (ALTERNATING-CURRENT CIRCUITS); VIBRATION DAMPING. Martin Greenspan

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Daphniphyllales

An order of flowering plants, division Magnoliophyta (Angiospermae), in the subclass Hamamelidae of the class Magnoliopsida (dicotyledons). The order consists of a single family with but one genus, *Daphniphyllum*, containing about 35 species. All are dioecious trees or shrubs native to eastern Asia and the Malay region. The plants produce a unique type of alkaloid (daphniphylline group); they often accumulate aluminum and sometimes produce iridoid compounds. The wood has vessels with scalariform perforations. The leaves are simple and entire, alternate or sometimes closely clustered at the ends of the branches. The flowers are small and inconspicuous, unisexual, regular, and hypogynous. Usually there are 2–6 sepals, or sometimes the sepals are absent; petals are lacking. The pollen is triaperturate, and the pistil has 2(–4) united carpels. *Daphniphyllum* has sometimes been included in the Euphorbiales of the subclass Rosidae, but structural details such as the very tiny embryo make it highly aberrant there. *See* HAMAMELIDAE; MAGNOLIOPSIDA; PLANT KINGDOM. T. M. Barkley

Dark current

An ambiguous term used in connection with both gaseous-discharge devices and photoelectric cells or tubes. In gaseous-conduction tubes it refers to the region of operation known as the Townsend discharge. The name is derived from the fact that photons produced in the gas do not play an important part in the production of ionization. The initial ionization arises from independent effects such as cosmic rays, radioactivity, or thermionic emission. When applied to photoelectric devices, the term applies to background current. This is current which may be present as the result of thermionic emission or other effects when there is no light incident on the photosensitive cathode. *See* ELECTRICAL CONDUCTION IN GASES; TOWNSEND DISCHARGE. Glenn H. Miller

Dark energy

An entity that comprises the majority of the energy of the universe and is responsible for the accelerating cosmic expansion. Its name derives from the inference that it is nonluminous. The nature of dark energy is speculative. Leading theories propose that dark energy is a static, cosmological constant

consisting of quantum zero-point energy, and a dynamical condensate of a new, low-mass particle. Alternatively, it has been proposed that the dark energy phenomena are due to a change in the form of gravitation on cosmological length scales. Confirmation of any one of these ideas would have a profound impact on physics. Dark energy appears to be distinct from dark matter, nonluminous particles which make up the majority of the mass of galaxies and galaxy clusters. Determining the nature of dark energy is widely regarded as one of the most important problems in physics and astronomy. *See* ACCELERATING UNIVERSE; COSMOLOGICAL CONSTANT; DARK MATTER.

Observational evidence. The empirical evidence for dark energy is indirect. Yet the evidence is manifest in what is, perhaps, the single most remarkable feature of the universe—the cosmic expansion. According to Albert Einstein's general theory of relativity, a cornerstone of the standard big bang cosmology, the rate of cosmic expansion is determined by the average energy density of all forms of matter and radiation in the universe. Measurements of the expansion, discovered by Edwin P. Hubble in 1929 and refined by the National Aeronautics and Space Administration's Hubble Space Telescope Key Project, therefore determine a critical energy density, $\rho_{\text{crit}} = 0.88\ (\pm 0.19) \times 10^{-9}\ \text{J/m}^3$ (68% confidence level), to which all forms of matter and radiation must sum. For comparison, the energy density of water is nearly 30 orders of magnitude larger. *See* BIG BANG THEORY; HUBBLE CONSTANT; RELATIVITY.

The first piece of evidence for the existence of dark energy arises from measurements of the gravitational mass density of galaxies and galaxy clusters. Diverse measurements accumulated over many decades, from Fritz Zwicky's 1933 observations to recent galaxy-redshift surveys, indicate that the fractional energy density in the form of atoms and even dark matter falls far short of the critical density, with $\Omega_m \equiv \rho_m/\rho_{\text{crit}} \approx 0.3$. The fraction in baryons, referring to all atoms or nuclear material, is a mere $\Omega_b \approx 0.04$, implying that the nature of ∼96% of the energy of the universe (26% dark matter and 70% dark energy) is unknown.

The second piece of evidence derives from measurements of the mean curvature of space. Just as mass and energy distort space-time, a mean curvature of space itself contributes an equivalent energy density. A space that is sufficiently curved so as to explain the dark energy must thereby possess a geometry in which initially parallel light rays traversing cosmological distances are distorted and eventually diverge. But detailed measurements of the cosmic microwave background radiation, most recently by NASA's *Wilkinson Microwave Anisotropy Probe*, have eliminated this possibility in confirming that the mean curvature of space is negligible. The ∼70% gap in energy must be due to something other than the curvature or dark and baryonic matter. *See* COSMIC BACKGROUND RADIATION; WILKINSON MICROWAVE ANISOTROPY PROBE.

The third piece of evidence, which provides the strongest clue to the nature of the dark energy, derives from the 1998 discovery of the accelerating cosmic expansion. Cosmological distance-redshift measurements of the scale of the universe as a function of look-back time reveal that the rate of expansion has been increasing during the last several billion years. In general relativity, the acceleration of a test particle due to the gravitational field of a fluid sphere is proportional to $\rho + 3p$ of the fluid, where ρ is the fluid density and p is the pressure. In the cosmic context, acceleration implies that the universe is under tension, like a fluid with a negative effective pressure. (An example is the stress parallel to a uniform electric or magnetic field.) Because no other cosmologically abundant forms of energy support a sufficiently negative pressure across cosmic scales, the dark energy must carry the cosmic pressure responsible for the acceleration. The strength of the pressure is characterized by an equation-of-state parameter, w, defined as the ratio of the homogeneous pressure to the energy density of the dark energy, $w \equiv p/\rho$. Formally, acceleration requires $w < -1/3(1 - \Omega_m)$, with current measurements giving $-1.2 \le w \le -0.8$ (95% confidence level).

These are the main lines of evidence for dark energy. Numerous other observations and measurements—including the age of the universe, the variation of distances with redshift, and distortions of the cosmic microwave background radiation by the accelerated expansion—are consistent with the interpretation that some form of dark energy with negative pressure is responsible for the majority of the energy of the cosmos.

Theoretical solutions. The leading hypothesis is that the dark energy is a cosmological constant (Λ), a uniform sea of positive energy and negative pressure with $w = -1$. A detailed comparison between theoretical predictions and observations finds that a cosmological constant supplying ∼70% of the critical energy density, or $\Omega_m \approx 0.3$, fits current data. In the absence of further clues, a cosmological constant provides the most economical description of dark energy.

The cosmological constant was introduced in 1917 by Einstein as a mathematical device in his ground-breaking application of general relativity to the problem of cosmology. A physical explanation of its origin was offered in 1967 by Yakov Zeldovich, who demonstrated that the vacuum of quantum-mechanical particles is equivalent to a cosmological constant. However, the predicted amplitude is at least 60 orders of magnitude too large—a mismatch regarded as one of the most enigmatic problems in theoretical physics. It is speculated that a complete understanding of the cosmological constant requires a theory of quantum gravity. *See* QUANTUM GRAVITATION.

The need to adjust or finely tune the cosmological constant to a value so many orders of magnitude below the predicted value has prompted speculation that perhaps $\Lambda = 0$ or that it does not contribute to the gravitational field equations, and some other mechanism is responsible for the accelerated expansion.

A widely studied possibility is that the dark energy is dynamical, consisting of a condensate of a new species of particles of very small mass. In the simplest version of this theory, these particles have mass $m \approx 10^{-68}$ kg and are identified as excitations of a cosmic field which oscillates with wavelength $\lambda \approx 10^{23}$ km, just within the size of the observable universe. This form of dark energy is also referred to as quintessence, as distinct from other fields and forms of matter or radiation. In contrast to a cosmological constant, the equation of state of quintessence is time varying, with $w > -1$, and its energy density and pressure fluctuate in response to gravitational field variations spanning length scales larger than λ. Through measurements of such time variation or fluctuations it is hoped to distinguish among competing theories of dark energy.

The theory of quintessence dark energy has two significant, although unproven, precedents. First, the dark energy phenomenon is very similar to primordial inflation, an epoch of accelerated expansion conjectured to have taken place in the early universe. According to inflation theory, the mechanism that drives the expansion is a cosmic field identical to quintessence except scaled up in mass and energy to match the conditions in the early universe. Second, the quintessence field is very similar to the axion particle, a dark matter candidate particle, originally proposed to resolve a fine-tuning problem in quantum chromodynamics. Despite these motivations for quintessence dark energy, any such theory also faces the challenge that it must explain why the new particle does not readily interact with known particles and mediate a new force, and why the new particles are so light compared to known particles. *See* INFLATIONARY UNIVERSE COSMOLOGY.

More speculative theories of dynamical dark energy with equation of state $w < -1$ have been proposed. One such example, called phantom dark energy for a similarity to so-called ghost fields in quantum field theory, is also a cosmic field but with very different properties from quintessence. The energy density of the phantom field grows with time, rather than decays as for quintessence. The viability of this theory has implications for quantum gravitational processes.

Numerous other theories of dynamical dark energy have been proposed. Most can be characterized simply by an equation of state of the quintessence ($w > -1$) or phantom ($w < -1$) type. In all cases, the magnitude of the cosmic acceleration is determined by the combination $-[1 + 3w(1 - \Omega_m)]/2$ so that, for equal matter density, quintessence gives the weakest acceleration and a phantom gives the strongest. Important clues to the nature of dark energy can be gained from measurements of the equation of state and any possible time variation.

Gravity. It has been proposed that the dark energy phenomena are due to a change in the form of gravitation on cosmological scales, as alternative to a cosmological constant or quintessence. Einstein's general relativity is well-tested within the solar system, and underpins the successes of the big bang cosmology. However, a modification of the gravitational field equations required to balance the expansion against the energy density and pressure of the cosmological fluid could obviate the need for the above-described theories of dark energy.

New theories of gravitation which introduce extra gravitational degrees of freedom under the rubric of scalar-tensor gravity, or which incorporate the effects of additional dimensions of space-time based on string theory or other fundamental theories of physics, are a subject of intense study. Most such theories proposed to explain the dark energy phenomena include a new cosmic field which is similar to quintessence but also modulates the strength of gravitation, thereby resulting in variability of Newton's constant, the amount of space-time curvature produced per unit mass or energy, and the degree to which the curvature produced by a massive body exceeds the sum of the curvatures due to its individual constituents. Most simple extensions of general relativity are heavily constrained by solar system observations in such a way that viable cosmological models are essentially indistinguishable from general relativity with a cosmological constant or quintessence dark energy. *See* SUPERSTRING THEORY.

Fate of the universe. The discovery of the cosmic acceleration suggests that the future evolution and fate of the universe will be determined by the nature of the dark energy. Previously, the main speculation was whether the big bang would terminate in a big crunch, as the universe collapsed under its own weight. Now a big chill appears as a likely fate: In a universe containing a cosmological constant or quintessence dark energy which maintains the inequality $-1 \le w < -1/3(1 - \Omega_m)$, the expansion will continue to accelerate forever. Only gravitationally bound structures today will survive into the distant future. Furthermore, the acceleration will prevent any communication across distances greater than $\sim 10^{23}$ km due to the appearance of a cosmological event horizon, and the universe will grow cold and rarefied. But in a universe containing phantom dark energy which maintains the inequality $w < -1$, the future would appear to be a big rip: the catastrophic end to the universe. In this case, the expansion would undergo a runaway process resulting in the gravitational disruption of clusters, galaxies, and then both gravitationally and nongravitationally bound objects on smaller and smaller scales. The Milky Way would be torn apart $\sim$50 million years before the big rip. The universe would be shredded to its fundamental constituents, unless quantum gravity effects intervene, and space-time itself would end $\sim$50 billion years in the future.

Prospective observations and experiments. The current challenge to cosmology is to tighten the evidence for dark energy, refine measurements of the cosmic expansion so as to narrow the bounds on the equation of state w, and continue to test the framework of general relativity and big bang cosmology. The main impact of the dark energy is on the rate of cosmic expansion, so that new clues to

the physics of dark energy may be obtained from measurements of distance, volume, and age versus redshift; the evolution and abundance of galaxies and clusters; and the distortion of the cosmic microwave background radiation. Numerous experiments and observatories dedicated to investigating dark energy are planned. Robert Caldwell

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Dark matter

Particles or objects that exert a gravitational force but do not emit any detectable light. Dark matter is the dominant form of matter in our Galaxy. Astronomers have detected the presence of dark matter through its gravitational effects and have shown that dark matter is not composed of ordinary atoms. Particle physicists have suggested several plausible candidates for dark matter; planned experiments are capable of detecting these new particles.

Astronomical evidence for dark matter. A variety of astronomical observations imply that dark matter is ubiquitous; it is detected in dwarf galaxies, in spiral galaxies, and in elliptical galaxies. It is the dominant form of matter in galaxy clusters and leaves clear signatures in the large-scale distribution of galaxies and in the microwave background. Astronomers infer the presence of matter through its gravitational effects. Since they have not been able to detect any light directly associated with this matter, they have labeled it "dark matter."

Dark matter is also sometimes called the "missing matter." This is a misnomer since astronomers detect the mass but they are unable to detect the light associated with the matter.

Galaxy motions in clusters. Many galaxies are bound together in galaxy clusters. These galaxy clusters contain hundreds and sometimes thousands of galaxies. These galaxies are moving with very high relative velocities. Typical velocities are roughly 1000 km/s (600 mi/s), and some galaxies that appear to be bound to the cluster move at velocities as high as 2000 km/s (1200 mi/s). Within an individual galaxy, stars and gas move much slower with typical velocities of only 200 km/s (120 mi/s).

As early as the 1930s, F. Zwicky recognized that clusters must be very massive to have a strong enough gravitational field to keep these high-velocity galaxies bound to the cluster. Zwicky noted that the inferred mass was much larger than the mass in stars. The result was so surprising that most of his colleagues ignored his conclusion. Current estimates suggest that stars comprise only 1% of the mass of the cluster (**Fig. 1**). *See* GALAXY, EXTERNAL.

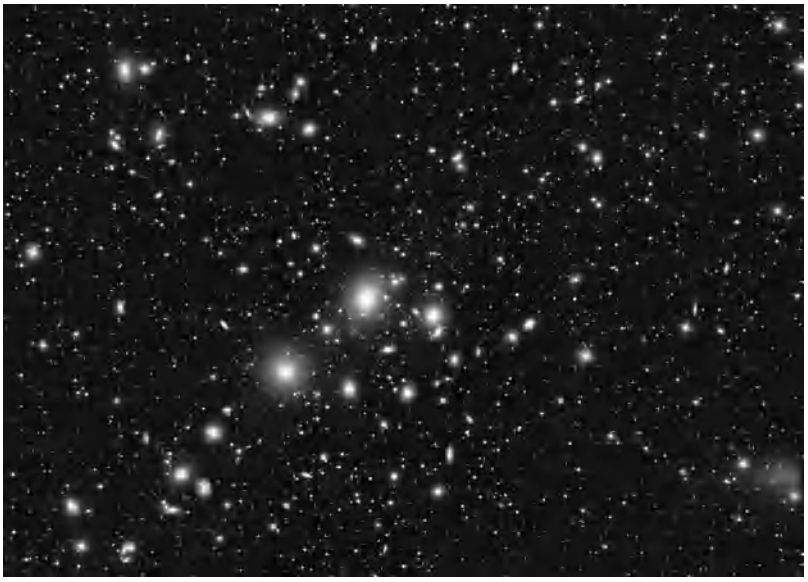

Fig. 1. Sloan Digital Sky Survey image of the Perseus galaxy cluster. Galaxies in the cluster are moving relative to each other at velocities of thousands of kilometers per second. Since the cluster appears to be gravitationally bound, this implies that the cluster has a mass of 10^{11} solar masses, roughly 100 times the mass contained in stars. (*Robert Lupton; SDSS Collaboration*)

Galaxy rotation curves. Most galaxies, including our own Milky Way, are spiral or disk galaxies, where cold atomic and molecular gas (primarily hydrogen) settles into a rotating disk. Much of the molecular gas is found in spiral arms. These spiral arms are active sites of star formation. *See* MILKY WAY GALAXY.

The gas and stars in disk galaxies move on nearly circular orbits. The Sun is moving on a nearly circular orbit around the center of the Milky Way Galaxy. To keep material moving on a circular orbit, centrifugal acceleration must balance gravitational acceleration. This balance can be expressed by the equation below, where V is the velocity of the gas or stars,

$$\frac{V^2}{R} = \frac{GM}{R^2}$$

R is the distance from the center of the galaxy, G is Newton's constant, and $M(R)$ is the mass contained within radius R. *See* CENTRIPETAL FORCE; CENTRIFUGAL FORCE.

Astronomers can measure the velocity of gas and stars in other galaxies through the Doppler effect. In the 1960s and 1970s, Vera Rubin pioneered optical observations of the motions of stars near the visual edges of galaxies. She found that the stars were moving rapidly. This large velocity implied the presence of hundreds of millions of solar masses of matter where little light was detected. Radio observations of the motions of neutral gas beyond the visual edge of spiral galaxies confirmed this remarkable result: Most of the mass in a galaxy is not in its disk of young stars, but rather in a halo composed of dark matter (**Fig. 2**). *See* DOPPLER EFFECT.

Hot gas in elliptical galaxies and clusters. X-ray satellites have detected copious amounts of hot gas in elliptical galaxies and in galaxy clusters. Since the thermal

Fig. 2. Sloan Digital Sky Survey image of galaxy M101. Measurements of the velocities in the disk of the galaxy imply the presence of enormous amounts of mass out toward the edge of the disk. (*Robert Lupton; SDSS Collaboration*)

pressure in this 1–10 million degree gas balances the gravitational field of the galaxy, measurements of the density and temperature profiles of the gas can be used to directly measure the mass distribution in these galaxies. These measurements reveal that the dark matter problem is ubiquitous: Stars can account for only a small fraction of the mass in elliptical galaxies. Observations of x-ray gas in clusters confirm Zwicky's inference that stars account for only a small fraction of the mass in a cluster (**Fig. 3**). *See* X-RAY ASTRONOMY.

Gravitational lensing. John Wheeler has remarked that "General relativity is simple: Matter tells space how to curve; the curvature of space tells matter how to move." Since mass curves space and distorts the path of light, images of galaxies that are behind rich clusters can be used to infer the distribution of mass in clusters. The masses in clusters distort background galaxies into arcs whose orientations are very sensitive to the distribution of matter in the cluster. Observations of gravitational lenses detect the presence of significant amounts of dark matter, and are consistent with the x-ray and optical observations of the same clusters (**Fig. 4**). *See* GRAVITATIONAL LENS; RELATIVITY.

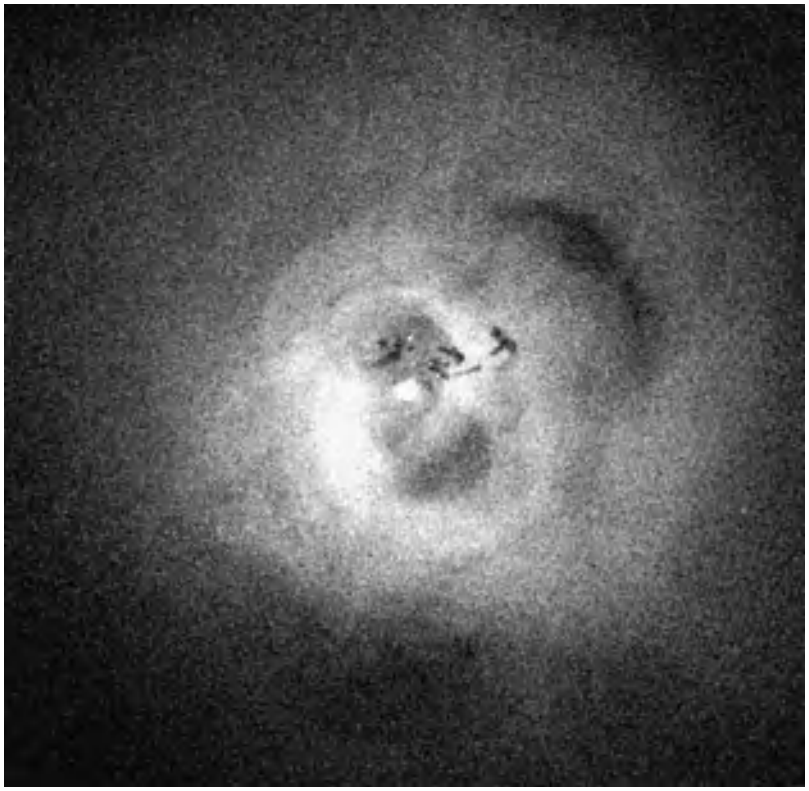

Fig. 3. *Chandra X-ray Observatory* image of hot gas in the Perseus galaxy cluster, indicating that stars account for only a small fraction of the mass of the cluster. (*NASA/Chandra X-ray Center/Institute of Astronomy/A. Fabian et al.*)

Fig. 4. *Hubble Space Telescope* image of a lensed background source in the form of an extended arc about an elliptical lensing galaxy. By measuring the distortions in the background, the mass of the foreground lensing galaxy can be inferred. (*Kavan Ratnatunga, Carnegie Mellon University; NASA*)

Novel nature of dark matter. The astronomical evidence discussed above shows that most of the mass in galaxies is not in the form of luminous stars and is consistent with a wide range of plausible candidates. Most astronomers assumed that the dark matter was similar to the "ordinary stuff" that makes up stars: protons, neutrons, and electrons. Cosmologists often refer to this "ordinary stuff" as baryonic matter. (Unfortunately, cosmologists use the term "baryonic matter" to refer to protons, neutrons, and electrons, while particle physicists, more correctly, refer to only protons and neutrons as baryonic matter.) Astronomers have searched for various possible forms of ordinary matter, including gas, dust, and low mass stars, but they have been unable to detect it. *See* BARYON.

Cosmological observations can measure the density of atoms in the early universe. These

observations imply that baryonic matter can account for only one-sixth of the mass in galaxies. These observations imply that most of the mass in galaxies is in the form of some novel and not yet identified form of matter.

Gas. Gas can be detected through either emission or absorption. Astronomers have measured the mass in atomic gas through observations of the 21-cm line and molecular gas through millimeter observations of carbon monoxide (CO) clouds. These observations suggest that only 1% of the mass of the Milky Way Galaxy is in cold gas in the galactic disk. *See* INTERSTELLAR MATTER.

X-ray satellites can detect 1–100 million degree gas through its thermal bremsstrahlung emission. These observations reveal that most of the baryonic matter in rich clusters is not in stars but in hot gas. In some rich clusters, there is ten times as much mass in hot gas as in stars. While significant for the evolution of the cluster, only 10–20% of the mass in a cluster is in hot gas. Dark matter accounts for most of the remaining 80–90%. *See* BREMSSTRAHLUNG.

Low-mass stars. While massive stars can be easily detected through their light, very low mass stars and planet-size objects are so dim that they can evade detection by optical telescopes. Astronomers, however, can detect these stars through their gravitational effects.

When a low-mass star passes in front of a distant star, it serves as a gravitational lens. This lensing brightens the image of the background star. Using large-area cameras capable of monitoring millions of stars, astronomers have searched for and detected these effects as low-mass stars pass in front of stars in the Large Magellanic Clouds and in front of the Milky Way bulge. While these observations have detected hundreds of low-mass stars through their gravitational effects, the number of events are consistent with models of the Milky Way Galaxy and imply that there are not enough low-mass stars to account for the dark matter.

Deuterium and helium. During the first minutes of the big bang, all of the neutrons in the universe combined with protons to make deuterium (one proton and one neutron). Two deuterium nuclei then collided to make helium (two protons and two neutrons). Most of the helium and deuterium in the universe was produced during the first few minutes of the big bang, and thus deuterium and helium can be said to be fossils from that time. The cosmological abundance of deuterium depends on the density of baryons (protons and neutrons) in the early universe: the more baryons, the lower relative abundance of deuterium. Astronomers use measurements of the ratio of the relative number of deuterium nuclei to provide a direct estimate of the density of atoms in the universe. *See* BIG BANG THEORY.

These observations imply that ordinary matter makes up only 4% of the energy density of the universe. These observations suggest that dark matter is not composed of protons and neutrons (or of nuclei made of protons and neutrons).

Microwave background fluctuations. Measurements of microwave background fluctuations are an important probe of the composition of the universe. The tiny variations in temperature seen by the *Wilkinson Microwave Anisotropy Probe* (and by ground and balloon-based microwave experiments) were produced during the first moments of the big bang. The statistical properties of these fluctuations depend on the composition of the universe: the more atoms, the smoother the shape of the fluctuations; the more matter, the higher the relative amplitude of small-scale fluctuations. The current observations confirm that ordinary matter makes up only 4% of the energy density of the universe and also imply that dark matter comprises 25% of the energy density of the universe *See* COSMIC BACKGROUND RADIATION; WILKINSON MICROWAVE ANISOTROPY PROBE.

Dark matter candidates. The big bang theory implies that the early universe was very hot and very dense, the perfect environment for creating particles through collisions. The densities and temperatures in the first microsecond of the big bang exceed the energies achieved in even the most powerful particle accelerators. Many cosmologists suspect that the dark matter is composed of some yet undiscovered fundamental particle that was produced in copious numbers during the first moments of the big bang.

Supersymmetry. The most popular dark matter candidate is the neutralino, a new particle posited by the theory of supersymmetry. Supersymmetry is an extension of the standard model of physics and is a vital element in almost all attempts to unify the forces of nature and in attempts to connect gravity and quantum mechanics.

Modern particle physics is based on the symmetries observed in nature. For example, all particles have antiparticles. The positron is the antiparticle of the electron. The antiproton is the antiparticle of proton. While seemingly bizarre, antimatter has not only been produced in the laboratory and detected in space, but now plays a major role in medical imaging (PET scans). *See* ANTIMATTER; MEDICAL IMAGING; SYMMETRY LAWS (PHYSICS).

Particle physicists have speculated on the existence of a new, not yet confirmed symmetry: supersymmetry. Supersymmetry implies that particles such as electrons and positrons have partners called selectrons and spositrons. These superparticles have the opposite statistical properties to ordinary matter. Selectrons behave like photons, while the supersymmetric partner of the photon, the photino, behaves quantum mechanically like an electron. This new symmetry has many aesthetic attractions and helps to explain the relative strength of the fundamental forces of nature. These new particles interact weakly with ordinary matter and have so far escaped detection. (They also may not exist.) *See* QUANTUM STATISTICS; SUPERSYMMETRY.

Supersymmetry implies the existence of a new stable particle: the neutralino. The neutralino would have been produced in abundance during the first moments of the big bang. During the first nanosecond of the big bang, there were as many neutralinos

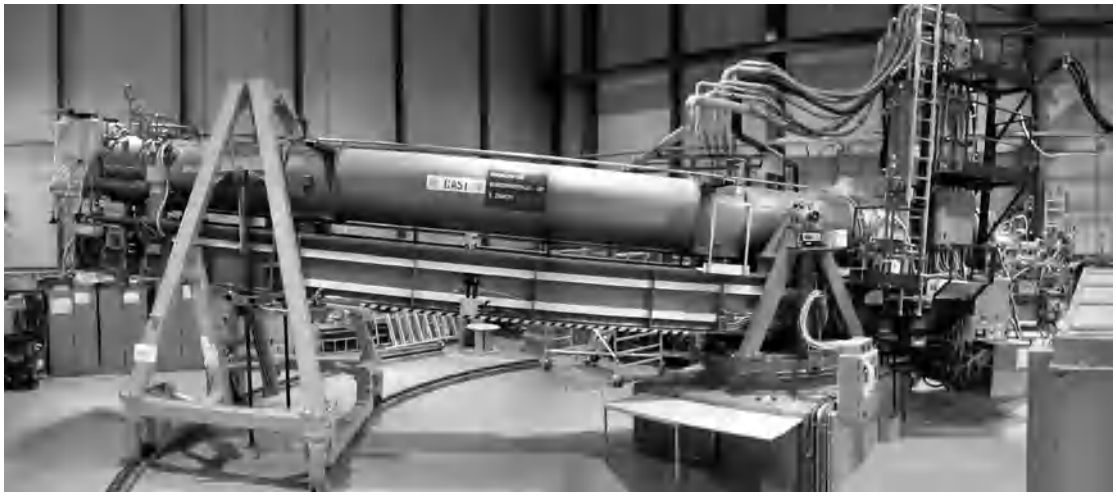

Fig. 5. CERN Axion Solar Telescope (CAST), which uses a 9-tesla, 10-m (33-ft) prototype dipole magnet for CERN's Large Hadron Collider, enhanced by the use of a focusing x-ray telescope, to search for axions from the Sun's core. (*CERN*)

as photons in the universe. While the number of neutralinos was reduced as they annihilated into ordinary matter, the current best estimates of their predicted residual abundance imply that they could be the dark matter. *See* WEAKLY INTERACTING MASSIVE PARTICLE (WIMP).

The primary scientific goal of the new Large Hadron Collider (LHC) at CERN in Switzerland, scheduled for start-up in 2007, is to detect the experimental signatures of supersymmetry. By colliding nuclei at velocities very close to the speed of light, the LHC aims to recreate some of the conditions in the big bang. The LHC may be able to provide direct evidence for the existence of supersymmetry and may even be able to produce dark matter particles in the laboratory. *See* PARTICLE ACCELERATOR.

Deep-underground experiments may also detect dark matter. If the neutralino is the dark matter, then hundreds of millions of these particles are streaming through our bodies every second. Because the neutralino interacts so weakly with ordinary matter, our bodies (and our detectors) are nearly transparent. These particles, however, do have rare weak interactions. In a kilogram detector, the dark matter particles may have a few collisions per day. These collisions are difficult to detect since each collision deposits only 10^{-3} (10^{-15} erg) of energy. Physicists have built a number of very sensitive low-background experiments capable of directly detecting neutralinos. Because the experiments need to be shielded from cosmic rays and other terrestrial sources of background signal, they must operate in deep-underground mines and tunnels.

Gamma-ray satellites may also be able to indirectly detect the neutralino. While the neutralinos interact weakly, occasionally one neutralino collides with another neutralino. This collision produces a shower of particles and radiation. Several on-going and planned experiments are looking for this dark matter annihilation signal. *See* GAMMA-RAY ASTRONOMY.

Axions. The axion is another hypothetical particle, which was invented to explain one of the symmetries seen in the strong nuclear interaction. If this explanation is correct, then axions would have been produced copiously during the first microsecond of the big bang. While axions interact very weakly with ordinary matter, they can be converted into ordinary photons in the presence of a very strong magnetic field. Experimentalists are using very strong fields to search for the axion (**Fig. 5**).

Black holes. If black holes were produced in large numbers during the first moments of the big bang, then they would be plausible candidates for the dark matter. Currently, there are no viable scenarios for producing large numbers of black holes in the early universe. However, since there are significant uncertainties in our understanding of physics during the first microsecond of the big bang, some astronomers consider black holes to be a viable dark matter candidate.

If the dark matter were composed of black holes and the black holes were significantly more massive than the Sun, then the black holes could be detected through their gravitational effects on galaxies and globular clusters, or through gravitational microlensing (discussed above in connection with the detection of low-mass stars). However, low-mass black holes could easily evade any current (and planned) observational searches. *See* BLACK HOLE.

Failure of general relativity. Many astronomers have speculated that the discrepancy between the mass seen in stars and gas and the mass inferred by applying general relativity (and Newtonian gravity, its low-velocity simplification) to astronomical observations is not the signature of some new exotic particle, but instead is the observational signature of the breakdown of general relativity. Classical Newtonian physics works well in daily life: Quantum mechanics is usually important only at very small scales and general relativity is only important on very large scales and near massive objects such as black holes and neutron stars. Perhaps the dark matter problem is the observational signature of the failure of general relativity. The discovery of dark energy

strengthens the motivation for considering these alternative models.

Modified Newtonian dynamics (MOND), developed by Mordehai Milgrom and Jacob Bekenstein, is the most carefully examined alternative gravitational theory. It has the advantage of explaining galaxy rotation curves without resorting to dark matter. However, MOND has difficulties fitting the data implying the existence of dark matter in clusters and is inconsistent with WMAP's observations of microwave background fluctuations. *See* COSMOLOGY; UNIVERSE.

David N. Spergel

Darmstadtium

The eighteenth of the synthetic transuranium elements, symbol Ds, atomic number 110. Darmstadtium should be a heavy homolog of the elements

1																	18
1 H	2											13	14	15	16	17	2 He
3 Li	4 Be											5 B	6 C	7 N	8 O	9 F	10 Ne
11 Na	12 Mg	3	4	5	6	7	8	9	10	11	12	13 Al	14 Si	15 P	16 S	17 Cl	18 Ar
19 K	20 Ca	21 Sc	22 Ti	23 V	24 Cr	25 Mn	26 Fe	27 Co	28 Ni	29 Cu	30 Zn	31 Ga	32 Ge	33 As	34 Se	35 Br	36 Kr
37 Rb	38 Sr	39 Y	40 Zr	41 Nb	42 Mo	43 Tc	44 Ru	45 Rh	46 Pd	47 Ag	48 Cd	49 In	50 Sn	51 Sb	52 Te	53 I	54 Xe
55 Cs	56 Ba	71 Lu	72 Hf	73 Ta	74 W	75 Re	76 Os	77 Ir	78 Pt	79 Au	80 Hg	81 Tl	82 Pb	83 Bi	84 Po	85 At	86 Rn
87 Fr	88 Ra	103 Lr	104 Rf	105 Db	106 Sg	107 Bh	108 Hs	109 Mt	110 Ds	111 Rg	112	113					

lanthanide series	57 La	58 Ce	59 Pr	60 Nd	61 Pm	62 Sm	63 Eu	64 Gd	65 Tb	66 Dy	67 Ho	68 Er	69 Tm	70 Yb
actinide series	89 Ac	90 Th	91 Pa	92 U	93 Np	94 Pu	95 Am	96 Cm	97 Bk	98 Cf	99 Es	100 Fm	101 Md	102 No

platinum, palladium, and nickel. It is the eighth element in the 6*d* shell. The various isotopes of darmstadtium have been reliably documented by the observation of at least four isotopes with half-lives in the range of milliseconds, a time too short for the application of chemical methods, verifying its position in the periodic system. *See* HALF-LIFE; NICKEL; PALLADIUM; PLATINUM; RADIOISOTOPE.

Searches for this element began in 1985. With the finding of the isotope ^{269}Ds in late 1994, its discovery became conclusive. Early experiments performed at FLNR (Flerov Laboratory for Nuclear Reactions), Dubna, Russia, GSI (Gesellschaft für Schwerionenforschung), Darmstadt, Germany, and LBNL (Lawrence Berkeley National Laboratory), Berkeley, California, failed to provide reliable evidence for a successful synthesis. In irradiations of uranium-235 and uranium-236 with beams of argon-40 at FLNR, a 10-ms spontaneous fission activity was observed in 1986 and assigned to the isotope ^{272}Ds. The method used for separation was a rotating wheel collecting the recoils combined with spontaneous fission detection. A production cross section of about 10^{-35} cm^2 was deduced. An irradiation of ^{235}U with ^{40}Ar at GSI in 1986 using the recoil separator technique reached the cross-section limit given by the FLNR experiment, but did not reveal any indication of darmstadtium formation. In 1991 at LBL bismuth-209 was irradiated by cobalt-59. A single sequence of signals observed after separation in a gas-filled separator was assigned to ^{267}Ds. But the correlation to known daughter isotopes of elements seaborgium to mendelevium could not be reliably established. Unsuccessful experiments to synthesize darmstadtium during 1984–1991 showed that production cross sections below 10^{-35} cm^2 should be made readily accessible by the equipment to be developed for future synthesis reactions. The technique was successfully applied at GSI for the synthesis of elements bohrium, hassium, and meitnerium. Local adaptations and improvements made this the standard technique in the 1990s. *See* ARGON; BOHRIUM; COBALT; HASSIUM; LEAD; MEITNERIUM; MENDELEVIUM; NUCLEAR REACTION; SEABORGIUM; URANIUM.

Separating heavy-ion products. Improved beam intensity and quality, improvement of the detection efficiency, and a new detector system allowing nearly complete chain reconstruction made the discovery of darmstadtium possible (**Fig. 1**). The total sensitivity for finding a new species was increased by a factor of 20. The efficiency to transport, select, and detect an atom produced as a fusion evaporation residue is 40%. The fusion evaporation residues observed in reactions of heavy-ion beams in thin targets (about 1 mg/cm^2) are generated with a momentum in direction of the beam of projectiles inducing the fusion reaction. Their recoil is well defined by momentum conservation and is large enough to allow the fusion evaporation residue to leave the target as a multiply charged heavy ion flying in the direction of the projectile beam with a velocity smaller than the velocities of any other reaction product. *See* ION.

At GSI the recoils are separated from the projectiles by a combination of electric and magnetic fields acting as a filter for a selected velocity. Two identical separation stages, each reducing the projectile beam by a factor of 10^5–10^6, isolate the fusion evaporation residue. About 3×10^{18} projectiles enter the velocity filter in a 10-day experiment, the time to detect one atom of an element produced with the cross section of 10^{-36} cm^2. The remaining 3×10^7 ions together with the hidden fusion evaporation residue leaving the filter are analyzed by a detector system. Once more the velocity of all ions is measured by a time-of-flight detector before they are implanted into a position-sensitive silicon detector of high energy resolution. The measurement of the deposited energy and the time of flight indicates the mass value for each of the ions. The fusion evaporation residue has the highest mass of all possible reaction products, and its implantation energy is defined by momentum conversation. Mass and energy of candidates for a fusion evaporation residue observed in their expected ranges are registered with the arrival time and the position. *See* ELECTRIC FIELD; ENERGY; ELECTROMAGNETIC FIELD; MASS.

The nuclei of the heaviest elements are radioactive. The main decay modes are alpha decay and spontaneous fission. Silicon detectors are best suited for detecting these decay modes. An alpha decay

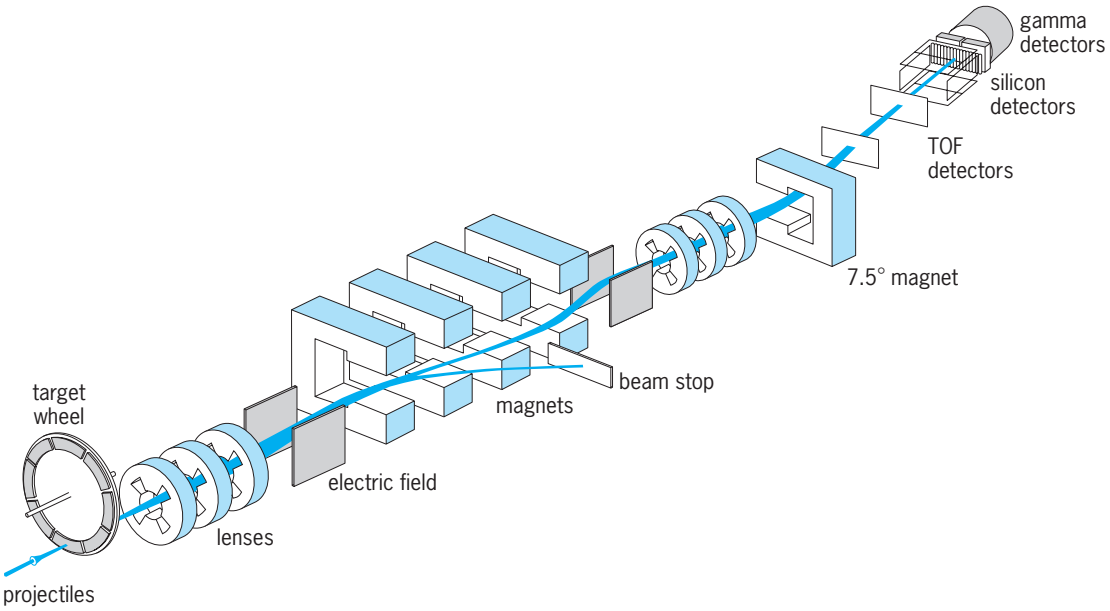

Fig. 1. A velocity filter is used as a separator for heavy-ion reaction products (SHIP). Fusion products travel at known velocities and are forced to follow the beam direction. Products are separated and identified by silicon detectors. TOF = time of flight.

does not change the position of the emitter in the bulk of the detector material within the accuracy of the position measurement. Correlated decays from one parent nucleus must have the same position coordinates. Spontaneous fission breaks the nucleus and ends its decay chain. As all correlation methods are restricted by the occurrence of random sequences, which are dependent on the rates in the windows defined for the different generations in the decay chain, the identification of an atom by its correlated decay chain becomes increasingly uncertain for long correlation times. Dividing the total measuring cycle into a time of irradiation with higher background rates and a measuring time between irradiations helps to reduce random correlations. Isotopes in the half-life range of 10^{-5} s to 1 min could be analyzed in GSI experiments in *x*-chains of up to six generations. For shorter decay chains, the maximum possible correlation times are in the few-seconds range.

The combination of recoil separation and single-decay-chain analysis allows detection of a new element from a single event observed. The method is fast (microseconds), efficient (40%), and sensitive (10^{-36} cm^2).

Decay chain. After an extensive search for the optimum bombarding energy on November 9, 1994, a decay chain was observed which proved the existence of the isotope ^{269}Ds. The isotope was produced in a fusion reaction of a nickel-62 projectile with a lead-208 target nucleus. The fused system, with an excitation energy of 13 MeV, cooled down by emitting one neutron and forming ^{269}Ds, which by sequential alpha decays transformed to hassium-265, seaborgium-261, rutherfordium-257, and nobelium-253. All of these daughter isotopes were already known, and four decay chains observed in the following 12 days corroborated without any doubt the discovery of the element. **Figure 2** shows the first decay chain observed, which ended in ^{257}Rf. The isotope ^{269}Ds has a half-life of 0.2 ms, and is produced with a cross section of about 3×10^{-36} cm^2. *See* ALPHA PARTICLES; NEUTRON; NUCLEAR STRUCTURE; RUTHERFORDIUM.

A second isotope, ^{271}Ds, was produced in a subsequent 12-day experiment by fusion of nickel-64 and lead-208. Nine atoms, with an excitation energy of 12 MeV, were produced, and by sequential alpha decay they transformed to the known isotopes

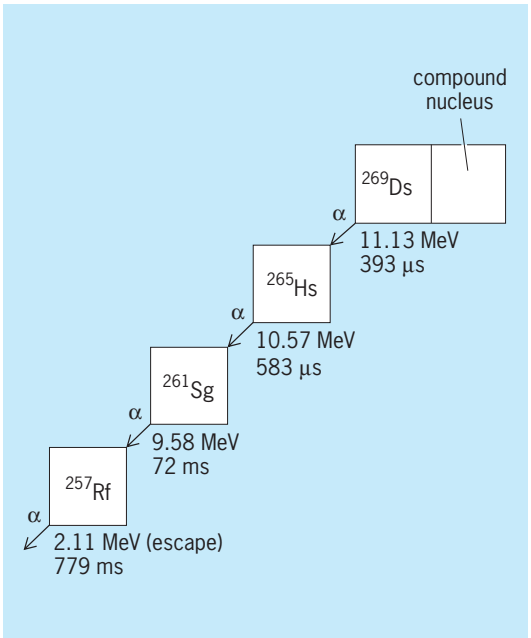

Fig. 2. Sequence of decay chains that document the discovery of darmstadtium. Numbers below boxes are alpha energies and correlation times. Darmstadtium is produced in the reaction ^{62}Ni + ^{208}Pb → ^{269}Ds + 1*n*.

hassium-267, seaborgium-263, rutherfordium-259, and nobelium-255. The half-life of ^{271}Ds is 1.1 ms, and its production cross section amounts to 1.5×10^{-35} cm^2. *See* ATOM.

In the decay chain of ^{277}112, the isotope ^{273}Ds was discovered in 1996. Its half-life is 0.1 ms, and the production cross section is 5×10^{-37} cm^2. The isotope ^{270}Ds with a half-life of 0.1 ms was synthesized at GSI-Darmstadt in the fusion of ^{207}Pb and ^{64}Ni. *See* ELEMENT 112. Peter Armbruster

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Dasheen

Common name for the plant *Colocasia esculenta*, including the variety *antiquorum* (taro). These plants are among the few edible members of the aroid family (Araceae). Native to southeastern Asia and Malaysia, the plants supply the people with their most important food. The edible corms (underground stems) support a cluster of large leaves 4–6 ft (1.2–1.8 m) long. A main dish in the Polynesian menu is poi, a thin, pasty gruel of taro starch, often fermented, which is frequently formed into cakes for baking or toasting. The raw corms are baked or boiled to eliminate an irritating substance present in the cells. *See* ARALES. Perry D. Strausbaugh; Earl L. Core

Dasycladales

An order of green algae (Chlorophyceae) in which the plant body (thallus) is composed of a nonseptate axis, attached by rhizoids and bearing whorls of branches. The walls utilize a partially crystalline mannan (polymer of mannose) as the skeletal component and usually are impregnated with aragonite, a form of calcium carbonate, causing these algae to be easily fossilized. Among green algae, only the Dasycladales have a fossil record sufficient to permit meaningful phylogenetic speculation. About 140 extinct genera have been described (**Fig. 1**) from limestones as old as the Ordovician, with three peaks of abundance and diversity—Carboniferous, Jurassic-Cretaceous, and Eocene. Only about 11 genera comprising 50 species are extant, with 3 of the genera extending from the Cretaceous. All are confined to warm marine waters except *Batophora*, which occurs as a relict in brackish sinkholes in New Mexico in addition to having a normal Caribbean distribution.

The life history is known for only a few genera, but the one that has been most thoroughly investigated—in *Acetabularia*—is unique. The stalk (**Fig. 2**), which may be as long as a few centimeters, bears a terminal whorl of sterile branches, which is superseded in the second growing season by another sterile whorl and in the third season by a fertile whorl. Prior to reproductive maturity, the thallus contains a single nucleus, which remains in a rhizoid and increases in diameter from about 2 to 150 micrometers. Increase in size is accompanied by development of a large nucleolar mass, which may be sausage-shaped or distributed among as many as 24 units. During the formation of the fertile whorl (whose component branches often cohere laterally to form a disk or cap), the nucleus undergoes profound but poorly understood changes, including reduction in size of the nucleus itself and in the size and number of nucleolar units. The residual nucleus

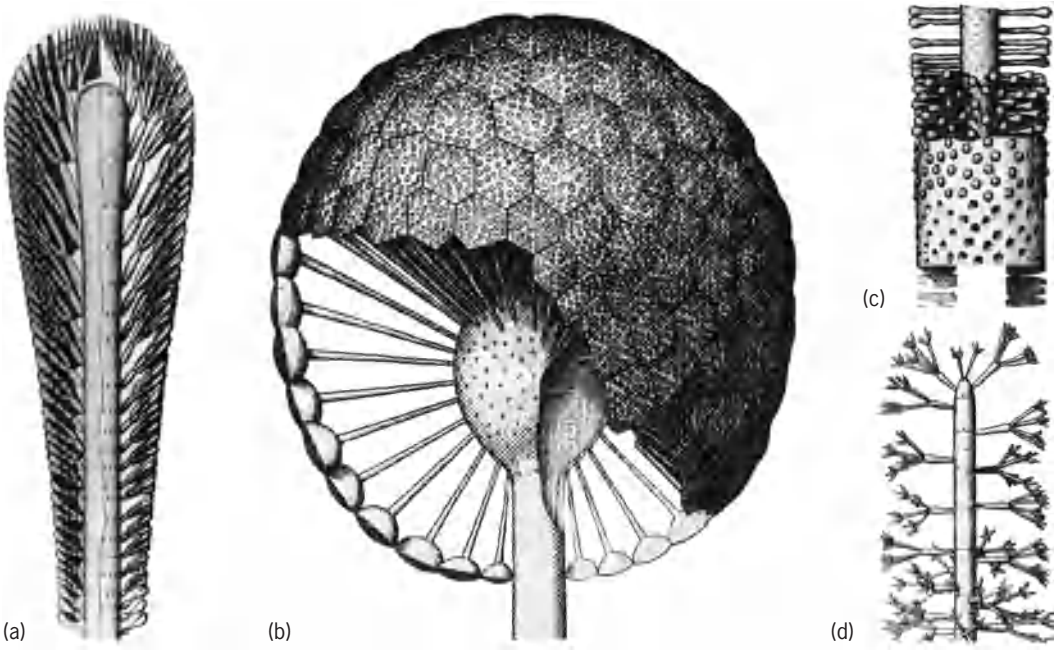

Fig. 1. Reconstructions of fossil Dasycladales. (*a*) *Palaeodasycladus* (Lower Jurassic, Italy); longitudinal section showing whorled branching to third order. (*b*) *Cyclocrinus* (Ordovician, Baltic); stalk terminating in head of unbranched laterals bearing reproductive cysts which cohere laterally to form a cortex. (*c*) *Rhabdoporella* (Upper Silurian, Sweden); single order of spirally arranged branches. (*d*) *Primicorallina* (Ordovician, New York); spiral branching to third order.

is still relatively large (about 40 μm in diameter), but has only one nucleolus. About 3 days following the completion of nuclear reduction, a gigantic intranuclear spindle is formed (to 120 μm long at late telophase). Although only two daughter nuclei are formed, chromosome counts support the belief that the division is meiotic.

A sequence of mitotic divisions results in the formation of numerous small secondary nuclei, which are carried by cytoplasmic streaming into the rays or branches (gametangia) that constitute the fertile cap. Each secondary nucleus together with surrounding protoplasm initiates the formation of an operculate cyst. Mitotic divisions prior to the onset of dormancy and immediately preceding germination result in numerous nuclei, which are incorporated in biflagellate isogametes. The cysts may become calcified and be liberated only when the parent plant disintegrates. Gametes from a single cyst are sexually incompatible. The quadriflagellate zygote swims briefly, then settles and develops into an irregular, prostrate, siphonous structure from which the primary axis arises.

Acetabularia (mermaid's wineglass) is the best-known genus of Dasycladales, partly because it occurs in the Mediterranean (as well as in all tropical seas), where its striking form gained the attention of early naturalists. Its extraordinary cytology and life history have made it a favorite subject of investigation by experimental morphologists, biochemists, enzymologists, biophysicists, and geneticists. The nucleus-containing rhizoid can be excised and the stalk cut into viable sections. Experiments that involve intraspecific and interspecific grafting of segments with or without a nucleus have provided important information concerning the nuclear control of morphogenesis.

In *Bornetella* (**Fig. 3**) and *Neomeris* the stalk is covered by lateral cohesion of the inflated apices of second-order branches. The thallus of *Bornetella* is spherical and only slightly calcified, with the gametangia enclosed within the continuous cortex. The thallus of *Neomeris* is cylindrical or clavate and heavily calcified, with the gametangia exposed at the base of the thallus by disintegration of the cortex. *Batophora* is exceptional in being completely uncalcified, its repeatedly divided branches remaining free. The thallus of *Cymopolia* is an articulate series of cylindrical, beadlike, calcified segments. This genus is exceptional in having dichotomous or irregular branching and coenocytic segments. Unlike *Acetabularia*, it produces gametangia continuously in uncalcified apical segments throughout growth.

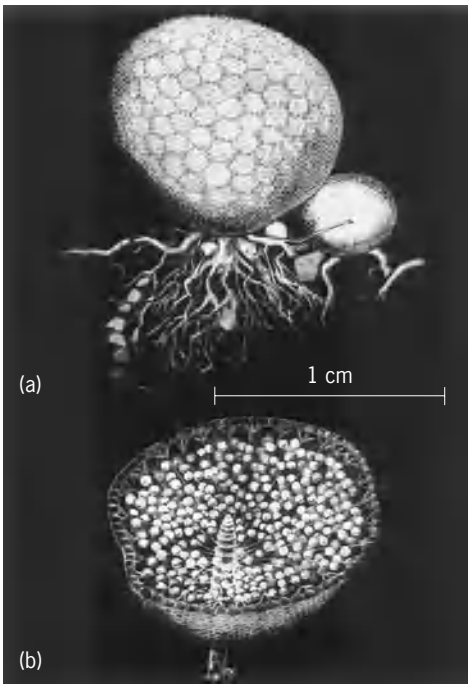

Fig. 3. ***Bornetella sphaerica*** **(extant). (*a*) Habit of thallus, showing sand grains trapped in rhizoids. (*b*) Oblique view into thallus, showing delicate primary branches bearing globose gametangia and whorls of secondary branches joined laterally to form a cortex.**

The extant Dasycladales are usually placed in a single family, Dasycladaceae, but those forms in which fertile and sterile whorls alternate or only one fertile whorl is produced may be segregated as the Acetabulariaceae (or Polyphysaceae). *See* ALGAE; CHLOROPHYCEAE.

Paul C. Silva; Richard L. Moe

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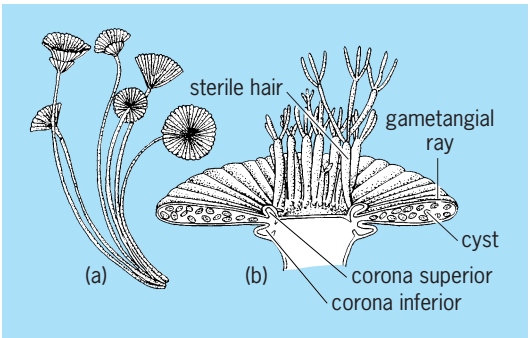

Fig. 2. ***Acetabularia*****. (*a*) Thallus of *A. crenulata*, showing stalk bearing fertile cap or disk. (*b*) Section of fertile disk of *A. acetabulum*. (*After G. M. Smith, Cryptogamic Botany, vol. 1, 2d ed., McGraw-Hill, 1955*)**

Data communications

The conveyance of information from a source to a destination. Data means the symbolic representation of information, generally in a digital (that is, discrete) form. (Analog information refers to information encoded according to a continuous physical parameter, such as the height or amplitude of a waveform, while digital information is encoded into a discrete set of some parameter, for example, a voltage level.) Usually, this digital information is composed of a sequence of binary digits (ones and zeros), called bits. The binary system is used because its simplicity is universally recognizable and because digital data have greater immunity to electronic noise than analog information and

allow flexible processing of the information. Groups of eight bits create a data byte or character. These characters make up the so-called alphabets (including alphabetic, numeric, and special symbols) which are used in data communications. The most common data sources and destinations are computers, computer peripherals, and other data devices [such as digital phones, cellular phones, personal digital assistants (PDAs), and digital imaging applications], and the data represent groups of characters to form text, hypertext (text organized according to topic rather than linear sequence), or multimedia information, including audio, graphics, animation, and video information. *See* BIT; COMPUTER; COMPUTER GRAPHICS; COMPUTER PERIPHERAL DEVICES; DIGITAL COMPUTER; ELECTRICAL NOISE; MULTIMEDIA TECHNOLOGY; NUMBERING SYSTEMS.

However, the term data communications more specifically deals with the methods by which this digital information is conveyed. The importance of data communications continues to grow. As society becomes increasingly information intensive and connected more applications require the best methods for conveying this information. Data communications may be accomplished through two principal functions, data transmission and data switching or routing.

Transmission and the data channel. Data transmission always involves at least three elements: a source of the information, a channel for the transmission of the information, and a destination for the information. In addition, sometimes the data are encoded. The codes can be used for error detection and correction, compression of the digital data, and so forth. This encoding and the corresponding decoding are specific to data in digital form. Depending on the application, the data can be transmitted point-to-point (for example, from a computer to its printer) or can be broadcast to many receiving points (similar to a television broadcast). *See* DATA COMPRESSION.

Transmission media. The communications channel is carried over a transmission medium. Such media can be wired, as in the cases of twisted-pair telephone wires, coaxial cables, or fiber-optic cables, or they can be wireless, where the transmission is not confined to any physical medium, such as in radio, satellite, or infrared optical transmission. Sometimes, even when the source of the information is digital, the transmission medium requires analog signaling, and modems (modulators-demodulators) are required to convert the digital signals to analog, and vice versa. For example, data communication between personal computers transmitted over telephone lines normally uses modems. Other transmission media (such as fiber systems) can support digital transmission directly. The choice of media frequently depends on the desired performance and cost criteria. *See* COAXIAL CABLE; COMMUNICATIONS CABLE; COMMUNICATIONS SATELLITE; MODEM; OPTICAL COMMUNICATIONS; OPTICAL FIBERS; RADIO; TELEPHONE SERVICE.

Parameters. The transmission is characterized by a number of parameters, including transmission directionality, whether or not the link is wired, whether bits are transmitted serially or in parallel, available capacity (bandwidth), data transmission rate (bits per second), channel error rate (the probability of errors per bit transmitted), channel reliability (how often the channel fails), and channel security (how easy it is to eavesdrop on the channel).

The directionality of the information can be either one-way (simplex communications) or two-way. Two-way communications can be either half-duplex (information goes both ways over the communications link, but not at the same time) or full-duplex (information goes both ways at the same time). The speed of transmission can vary greatly, depending on the application, from a few bits per second (for applications such as alarm monitoring or point-of-sale terminals) to advanced fiber-optic transmission systems, where many billions of bits per second can be transmitted along a single optical fiber. (For comparison, a typical 400-page book, entirely text, contains approximately 15–25 million bits.)

The data channel can be a serial channel, in which the bits are transmitted one after another across a single physical connection; or a parallel channel, in which many bits are transmitted simultaneously (for instance, over parallel wires). Fiber-optic communications, for example, is usually serial, whereas the backplane of a computer, where data are exchanged between memory and processing elements, is usually parallel. Generally, parallel channels are used for short-distance links (less than 300 ft or 100 m), whereas serial links are used for larger distances and high data rates.

At low data rates (less than a few hundred kilobits per second) communications channels are typically dedicated, whereas at higher data rates, because of the cost of high-speed transmitters and receivers, shared channels are used by multiplexing the data streams. For example, two independent data streams with constant data rates of 10 megabits per second (Mb/s) could use a shared channel having a data-rate capability of 20 Mb/s. The multiplexing system would select one bit from each of the two channels to time-multiplex the data together. Similarly, at the destination the channels could be demultiplexed to restore the independent 10-Mb/s channels (**Fig. 1**). Here, the available information-carrying capacity of the channel (20 Mb/s) is shared by the two users through multiplexing. There are many different types of multiplexing used to share the channels; for example, there can be multiplexing in frequency, time, coding, or amplitude. *See* MULTIPLEXING AND MULTIPLE ACCESS.

In choosing the technology for transmission, the appropriate parameters depend on the application requirements. For example, the link between a personal computer and a printer requires lower performance (and lower cost) than that connecting two high-performance computers.

Switching. In some applications, the data communications link can be a permanent, dedicated connection, for example, from a computer to its printer. However, in many cases, the source, the destination, and the path taken by the data may vary; thus switching is required.

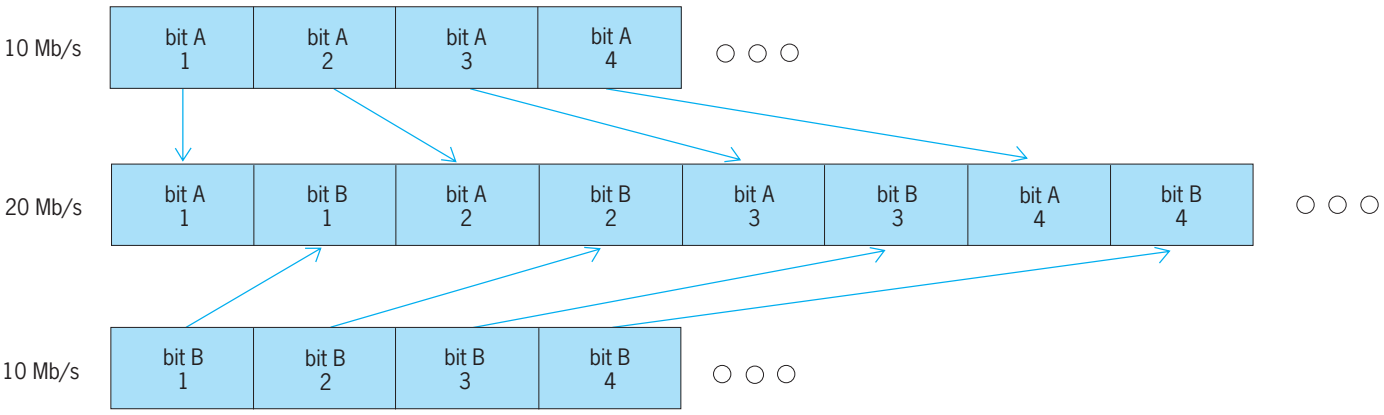

Fig. 1. Bit-stream multiplexing.

Types. The two primary types of switching employed in data communications are circuit switching and packet switching. In circuit-switched data communications, an end-to-end connection is established prior to the actual transmission of the data and the communications channel is open (whether or not it is in use) until the connection is removed. Since the full information-carrying capacity of the channel (the bandwidth) is available to the user, no additional connections along that channel are allowed (that is, new requests for the use of the channel are blocked) as long as that connection is in place.

A packet is a group of data bytes which represents a specific information unit with a known beginning and end. The packet can be formed from either a fixed or variable number of bytes. Some of these bytes represent the information payload, while the rest represent the header, which contains address information to be used in routing the packet. The packet is somewhat analogous to an envelope. The information payload is the content of the envelope, whereas the header is like the address on the envelope.

In packet switching, unlike circuit switching, the packets are sent only when information transmission is required. The channel is not used when there is no information to be sent. Sharing the channel capacity through multiplexing is natural for packet-switched systems. Furthermore, packet switching enables users to access the full capacity of the communications links, even when the links are already in use; that is, the packet switches allow for temporary loading of the network beyond the transmission capacity of the channel. This information overload is stored (buffered) in the packet switches and sent on when the channel becomes available. (Thus packet switching is sometimes referred to as store-and-forward switching.) Of course, such network overloads must be temporary.

A characteristic of data communications traffic (the amount of data being transmitted through a link) is that it is often bursty. That is, the traffic is not constant in time, but has intense periods and very light periods. Thus, packet switching is well suited to this type of traffic. An important characteristic of the information being sent is the allowable delay and loss characteristics. Voice-telephone traffic and video traffic, for example, require constant bit rates without variable delay. If a several-second pause in the transmission were to occur, it would be disturbing to a phone call or video transmission. In contrast, for most data traffic a few-second delay may not even be noticed (for example, from a point-of-sale charge-card terminal to a bank). However, data signals require very low losses (a single bit error can cause retransmission of a whole packet), whereas a few bits lost for audio or video signals may be easily tolerated.

Although circuit switching is widely used in the voice-telephone network, where calls are established for holding times typically on the order of minutes, modern data communication networks are largely structured around packet-switching technologies. Packet switching provides flexibility between various types of equipment with inherently different data rates, as well as efficient sharing of the transmission capacity between bursty data sources.

Protocols. In order to transfer information from a sender to a receiver, a common physical transmission protocol must be used. Protocols can range from very simple to quite complex. For example, a protocol widely used for low-data-rate serial data transmission from a computer to and from its low-speed peripherals is defined by a common set of electrical specifications referred to as RS-232-C. (Actually, the RS-232-C protocol uses analog signaling between digital devices.) Today the most widely used data link protocol is the Ethernet protocol. In the local-area network (LAN) it is used between 10 and 100 Mb/s, and between Ethernet switches it is used at either 1 gigabit per second (Gb/s) or 10 Gb/s. More than 95% of all data packets originate on Ethernets. The most widely used protocol for data communications over a vast range of technologies is the Internetwork Protocol (also known as the Internet Protocol or IP).

Existence of connections. Data communications can be connection-oriented or connectionless. In connection-oriented systems, each time that information is required, a connection is established and the information is passed over the connection path. In connectionless systems, the header information in the data packet is used by each of the packet switches along the path to route the packet to its destination.

Packet-switched networks. A network is composed of a number of communication links working together. The complexity of the data communications can depend on the interconnection and network size. At the shortest distances, communications occur between integrated circuits (chip-to-chip interconnection). This is followed by interconnections between circuit boards containing the chips (board-level interconnection). Groups of boards are interconnected (backplane interconnection) to form computers and other digital processing systems, and computers are interconnected to form local-area networks (LANs). Local-area networks are interconnected to form metropolitan-area networks, and these are interconnected to form wide-area networks (WANs). Finally, the wide-area networks are interconnected to form national-scale networks, which are interconnected to form global networks spanning the world. *See* LOCAL-AREA NETWORKS; WIDE-AREA NETWORKS.

Protocol layering. As network complexity grows, a hierarchy of layers is used to better manage and control the network. The layer definitions and the protocol definitions are portions of the network architecture. The International Standards Organization developed a widely used model for these layers, called the Open Systems Interconnect (OSI) model (**Fig. 2**). The essential feature of the OSI model is that it reduces protocol complexity by breaking the protocol into smaller independent functional units which operate in conjunction with matched functional units at a peer-to-peer level. Each layer performs functions for the next higher layer by building on the functions provided by the layer below. The advantage of performing communications based on this model is that the protocol at each layer can be defined independent from the other layers. Standard protocols at each layer can be applied, enabling internetworking.

At the lowest layer, there is physical transmission, which is concerned with transmitting raw data bits. Important issues at this layer include the transmission data rate, the bit error rate, and the delay through the links. Above this physical layer are virtual communications layers, designed to simplify the management of complex functions. They are virtual in the sense that there is no additional physical connection at any layer above the physical transmission. The data-link layer takes the physical transmission of bits and makes it appear to the higher layers that the communications are error free. This may be accomplished by defining some bytes in the information as check bytes, to detect transmission errors. Building on this, the network layer has the job of converting the information into packets and determining the routing of the packets. This layer has to deal with the flow of the traffic in the network and possible network congestion.

The transport layer accepts information from the user layers, apportions it into smaller units for transmission through the network, and is responsible for regrouping the units to ensure that they arrive at the destination in the correct order and without errors. The transport layer is responsible for end-to-end flow control. For example, in the communications between a computer and a printer, the transport layer controls the flow of data out of the computer so that the printer is not overwhelmed by it. This layer also determines what type of service to provide to the higher layers. The data-communications-networking functionality is largely confined to these first four layers in the network hierarchy. The highest layers are related to the end-user applications.

Ethernet protocol. Some protocols take explicit advantage of the bursty nature of data communications. For example, the Ethernet protocol (carrier sense multiple access and collision detect, or CSMA-CD) is based on a common bus (ether) in which every transmission is broadcast to all users. The Ethernet protocol works as follows: When a host computer wants to transmit a packet, it first listens to the physical medium (the ether) to see if it is currently being used. If not, it transmits its data packet. If the ether is busy, it waits a specific time interval, then tries to transmit again. This same protocol can be used for all the hosts on the ether. The host computers

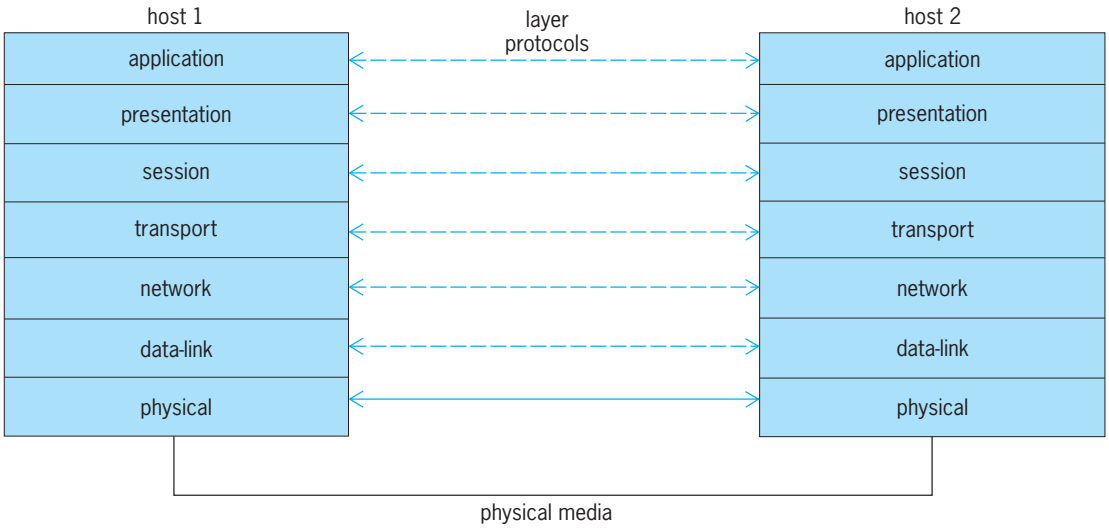

Fig. 2. Open Systems Interconnect (OSI) model for network protocols.

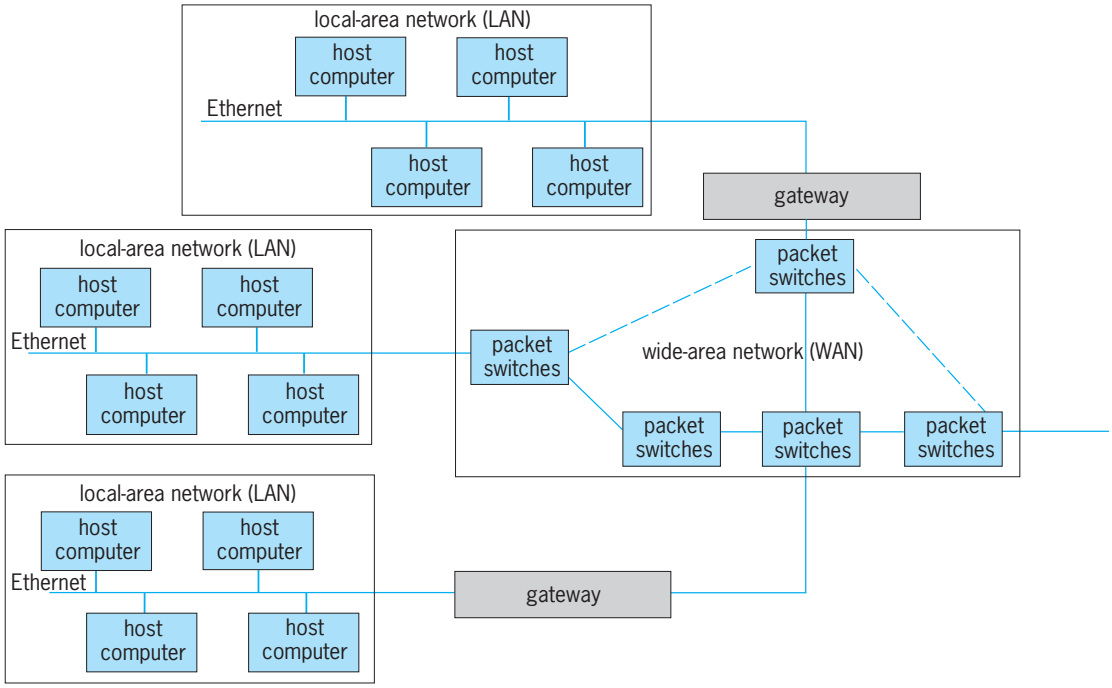

Fig. 3. Data communications network.

must be able to detect packet collisions caused by two hosts simultaneously transmitting on the ether. Collision recovery is accomplished by aborting the transmission, waiting a random length of time, and then trying again.

While the Ethernet (CSMA-CD) would not make sense for telephone traffic, it is very efficient for bursty data communications within a LAN. This type of system is feasible only if the packet propagation time (the time to travel from the source to the destination) is short relative to the packet transmission time (the time required to send the packet). Thus, this type of protocol is well suited to LANs, but would not be useful in large-scale networks (for example, over WANs or satellites), where the propagation time can be large relative to the transmission time because of the great distances involved. However, the original Ethernet protocol has been extended to a switched Ethernet protocol in which there is bandwidth reservation rather than packet collision detection as described above. This protocol can be used at data rates from 10 and 100 Mb/s to more than 10 Gb/s and over large distances.

Network interconnection. To form wide-area networks, multiple LANs are usually interconnected by using gateways or bridges, which provide communications paths so that data may be exchanged between networks (**Fig. 3**). To achieve integration of these subnetworks into a larger framework of networks, these gateways must support internetworking protocols. The most common internetworking protocol is based on a system known as TCP/IP (transmission control protocol/internetwork protocol). The IP protocol provides the networking protocol portion of the OSI model and TCP provides the transport layer.

Internet. The Internet is an immense data network of networks. It has grown from an experimental network called the ARPAnet, originally operated by the U.S. Defense Advanced Research Projects Agency and using the TCP/IP protocol, into the world's largest network of networks. The Internet now supports a huge variety of protocols in addition to the TCP/IP protocol (many of them using the TCP/IP protocol). The Internet has had pervasive impact on all aspects of communications. It has revolutionized communications for commerce, science, and art, and has had a large-scale impact on communications in daily life. It has given rise to many other protocols, including hypertext transmission protocol (HTTP) and the World Wide Web (WWW), and the many familiar browsers and higher-level applications.

Addressing in the Internet is accomplished through a specific protocol version. While the version known as IPv4 is widely used, an evolution is taking place to greatly extend the original addressing scheme. IPv6 is a new version of IP which is designed to be an evolutionary step from IPv4. IPv6 is designed to run well on high-performance networks (such as Gigabit Ethernet, OC-12, and ATM) and at the same time still be efficient for low-bandwidth networks (such as wireless networks). In addition, it provides a platform for new Internet functionality that will be required in the near future. IPv6 includes a transition mechanism which is designed to allow users to adopt and deploy IPv6 in a highly diffuse fashion and to provide direct interoperability between IPv4 an IPv6. The IPv6 transition allows users to upgrade their hosts to IPv6, and the network operators to deploy IPv6 in routers, with very little coordination between the two.

The Internet enables the integration and convergence of voice, data, and video transmission, and it is commonly used for Voice over IP (VoIP) and streaming video transmission as well as for its originally

intended computer data networking applications. The Internet also supports many higher-layer applications such as the World Wide Web. There is no single coordinating or operating group for the Internet, but many of the developments and enhancements to the Internet originate from the Internet Engineering Task Force (IETF), a large international group of network architects, designers, equipment vendors, and carriers, who recommend evolutionary changes to the Internet. Use of the Internet is the largest, most pervasive electronic connectivity in human history—it is the networking parallel of the development of microelectronics. *See* INTERNET; VOICE OVER IP; WORLD WIDE WEB.

Other networking protocols. One advanced protocol, frame relay, supports packets (frames) of variable length, from a few bytes to around 10,000 bytes. Frame-relay functions are connection-oriented; that is, frames of individual virtual connections are transferred by means of address information in the control part of the frame. Error correction is implemented by retransmission of faulty frames end to end. The network switches only frames for which no errors were found, and discards all faulty frames. Frame relay is typically used for applications at bit rates up to about a few megabits per second. Frame relay, while suitable for a range of data communications applications, has limited applicability to multimedia data involving time-sensitive information (such as audio or video), which must be delivered without large, variable delays.

Asynchronous transfer mode. Communication between powerful desktop computing platforms generates large volumes of data communications traffic, which requires high transmission speeds. Furthermore, demands are placed on LANs and other networks for multimedia traffic, and include variable-bit-rate data as well as continuous-bit-rate voice and video. At these higher data rates, packet-switching protocols based on the asynchronous transfer mode (ATM) are used. Variable-length packets can be difficult and slow to forward through networks, and may cause performance limitations to continuous-bit-rate services. ATM addresses this problem through a packet-switching protocol where fixed-length, 53-byte packets (cells) are used. The information payload is 48 bytes, with 5 bytes for header information. ATM may be used with systems at data rates ranging from low (around 1.5 Mb/s) to very high (up to 2.5 and 10 Gb/s). ATM may be suitable for both bursty and continuous-bit-rate traffic. Bursty traffic is assigned a lower priority so that it takes advantage of the statistical multiplexing possible for such traffic, while reasonable performance is guaranteed for continuous-bit-rate services which typically have end-to-end delay requirements. While ATM is widely used in the public switched telephone network and some data networks, due to its complexity, it is being replaced by Switched Ethernet.

Switched Ethernet. As described above, Ethernet began as a LAN technology, but it has continued to evolve and is also used as a metro- and wide-area-networking technology. These capabilities began with the development of 1-Gb/s Ethernet capability, and 10-Gb/s Ethernet has become prevalent. The advantages of Ethernet are its low cost due to its widespread use and large market in the LAN and enterprise areas, and its relative simplicity.

Multiprotocol label switching. Another set of protocols is called Multiprotocol Label Switching (MPLS), which is an IETF initiative to integrate layer 2 information (bandwidth, utilization, and so forth) about network links to layer 3 and to simplify and improve the performance at layer 3 (the IP layer). Simply put, when a packet enters an MPLS network, label edge routers give it a label. These labels contain not only information on how to route the packets but also information about the types of service that are being requested and the protocols that are being used.

MPLS enables routing of the information based on this label information, so that different qualities of service may be supported. MPLS gives network operators a great deal of flexibility to divert and route traffic around link failures, congestion, and bottlenecks. It essentially adds "superhighways" and can help address network congestion at specific routers. There are a number of generalizations and variants of MPLS as well.

Wireless protocols. The need for mobility in data communications has given rise to the development of a large number of wireless networking protocols. Some of the most common are the IEEE 802.11 suite of protocols. These layer 3 and 4 protocols address the unique requirements of the wireless environment. *See* ELECTRICAL COMMUNICATIONS; INTEGRATED SERVICES DIGITAL NETWORK (ISDN); PACKET SWITCHING; SWITCHING SYSTEMS (COMMUNICATIONS); TELEPROCESSING; WIRELESS FIDELITY (WI-FI).

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Data compression

The process of transforming information from one representation to another smaller representation from which the original, or a close approximation to it, can be recovered. The compression and decompression processes are often referred to as encoding and decoding. Data compression has important applications in the areas of data storage and data transmission. Besides compression savings, other parameters of concern include encoding and decoding speeds and workspace requirements, the ability to access and decode partial files, and error generation and propagation.

Basic Principles

The data compression process is said to be lossless if the recovered data are assured to be identical to the source; otherwise the compression process is said to be lossy. Lossless compression techniques are requisite for applications involving textual data. Other applications, such as those involving voice and image data, may be sufficiently flexible to allow controlled degradation in the data.

Data compression techniques are characterized by the use of an appropriate data model, which selects the elements of the source on which to focus; data coding, which maps source elements to output elements; and data structures, which enable efficient implementation.

Source data. The source data typically are presented in a format in which each datum is represented by a fixed number of bits. For textual data, each character is typically represented in eight bits using, for example, the extended American Standard Code for Information Interchange (ASCII) mapping. For speech, each datum consists of a real-valued sample of the waveform expressed to a set precision. For images, each picture element or pixel typically is represented by 1 bit (for black-white), 8 bits (for gray scale), or 24 bits (for color images, 8 bits intensity for each of three primary colors).

Models. Information theory dictates that, for efficiency, fewer bits be used for common events than for rare events. Compression techniques are based on using an appropriate model for the source data in which defined elements (source values in general or within a particular context) are not all equally likely. The encoder and the decoder must agree on an identical model. The encoder determines the relevant elements under the agreed model and outputs sufficient information to enable the decoder to determine these elements. *See* INFORMATION THEORY.

A static model is one in which the choice of elements and their assumed distribution is invariant. For example, the letter "e" might always be assumed to be the most likely character to occur. A static model can be predetermined with resulting unpredictable compression effect, or it can be built by the encoder by previewing the entire source data and determining element frequencies. This latter process suffers from the necessity of the encoder making two passes over the data, one pass for determining frequencies and another for encoding, and also from the delay imposed on producing any output until after the entire source has been scanned. In addition, the relevant aspects of the determined frequency distribution must be sent to the decoder, which would otherwise have no means of determining it. The benefits of using a static model include the ability to decode without necessarily starting at the beginning of the compressed data.

An alternative dynamic or adaptive model assumes an initial choice of elements and distribution and, based on the beginning part of the source stream that has been processed prior to the datum presently under consideration, progressively modifies the model so that the encoding is optimal for data distributed similarly to recent observations. Some techniques may weight recently encountered data more heavily. Dynamic algorithms have the benefit of being able to adapt to changes in the ensemble characteristics. Most important, however, is the fact that the source is considered serially and output is produced directly without the necessity of previewing the entire source.

In a simple statistical model, frequencies of values (characters, strings, or pixels) determine the mapping. In the more general context model, the mapping is determined by the occurrence of elements, each consisting of a value which has other particular adjacent values. For example, in English text, although generally "u" is only moderately likely to appear as the "next" character, if the immediately preceding character is a "q" then "u" would be overwhelmingly likely to appear next.

In a sequential model, the source is considered in order of its (one-dimensional) presentation. In a hierarchical model, the source is recursively partitioned into smaller subparts. The quadtree representation, which recursively splits a square into four subsquares, exemplifies the use of this model.

Coding techniques. The use of a model determines the intended sequence of values. (For example, if a dictionary model is used for text compression, the output may be indices into a hash table. If a gray image is compressed, the output may correspond to gray-scale values, typically in the range 0–255.) An additional mapping via one coding technique or a combination of coding techniques is used to determine the actual output. Several commonly used data coding techniques will be described. The decoder is made aware of which coding technique has been employed either by prearrangement or by an inserted codeword.

Dictionary substitution. In using this technique, a repeated sequence of characters is replaced with a reference to an earlier occurrence. A table to possible values for reference is maintained; a value from this table is presented as an index into the table. In ordered table indexing, the table of possible values is ordered by some predetermined criteria. Actual frequency can be used, necessitating either a prescan or the maintenance of a corresponding dynamic frequency table. Alternatively, a strategy for self-organizing a list, such as transpose or move-to-front, may be used. The method of static dictionary substitution is useful for representing alphanumeric representations of information, such as gender or month. In a Lempel-Ziv (or dynamic dictionary) approach, as additional source characters are scanned, some table entries may be deleted and appropriate candidates may be considered for addition to the table. The variations on this method differ in terms of how the references to earlier occurrences are presented and how the deletions and additions to the table are selected.

A typical Lempel-Ziv procedure is as follows. The table is initialized to contain the set of singleton characters. Iteratively the longest string s in the table that matches the source is determined, and the corresponding table index is output. If the following

nonmatched source character is *c*, then the concatenation of *s* followed by *c* is added to the table if space remains. When the table becomes full, this method reverts to static dictionary substitution. Alternatively, after the table becomes full and the saving from compression is observed to degrade below a trigger level, the table is reinitialized. Another approach maintains a sliding window in which string references may occur. For example, a longest match of the to-be-compressed source is sought within the immediately proceding 4095 characters of the already-compressed source.

Run-length encoding. With this approach, a sequence of values is presented as a sequence of pairs (c,v), where c is the count of the number of values in a row that have the same value v or are within a predefined margin of allowable error.

Predictive coding. For this technique, a sample value is presented as the error term formed by the difference between the sample and its prediction. In difference mapping, it is expected that the next value is close to the current value, and so the predicted value is defined to be equal to the previous sample. In a linear predictive model, the expectation of the next value is the extrapolation of the present and previous values. Thus, if the previous value was 200 and the current value is 220, the next value will be 240. There are many other predictive models, both one- and two-dimensional, in use.

Huffman coding. The essence of this coding technique is that a frequency or probability distribution is associated with the set of candidate values. A code tree is constructed in such a manner as to minimize the expected weighted path length of the tree. A value is presented as a path from the root of the code tree to a leaf, expressed as a sequence of 0's and 1's denoting left and right tree branches.

Arithmetic coding. This technique is based on the concept that associated with the set of candidate values is a frequency or probability distribution from which can be determined a cumulative frequency distribution and, ultimately, associated subintervals of the unit interval. For example, if the only characters are *X* with probability 20%, *Y* with probability 70%, and end-of-string (EOS) with probability 10%, then *X* occupies the interval 0 to 0.2, *Y* occupies 0.2 to 0.9, and EOS occupies 0.9 to 1. Accordingly, associated with a sequence of values will be an increasingly smaller subinterval of the unit interval. A sequence of values is presented as a real number within that ultimate small subinterval. Continuing with the earlier example, the string *YX* would be associated with the interval 0.326 to 0.34 because the first character (*Y*) focuses on the interval 0.2 to 0.9, the second character (*X*) focuses on the first 20% of that interval (0.2 to 0.34), and the string end focuses on the last 10% of that. Similarly, the string *XY* would be associated with the interval 0.166 to 0.18. The code is the final real number to whatever precision is required to distinguish it from other contenders.

Variable-length representation of integers. A number of different coding techniques have been developed that, depending on the relative expected frequency of integers, can present the values of the integers in compact form. Such codes include Elias codes, Fibonacci codes, and start-step-stop codes.

Scalar quantization. In this technique, a value is presented (in approximation) by the closest, in some mathematical sense, of a predefined set of allowable values. In some cases, the allowable values may be nonuniformly spaced. For example, a real number may be presented rounded to the nearest three decimal places or truncated to five significant digits.

Vector quantization. In this method, a sequence of k (some fixed number) values is presented as resembling the template (from among the choices available in a given codebook) that minimizes a distortion measure. Under this approach, the encoder must output information sufficient for the decoder to be aware of the contents of the codebook as well as the selection.

Transform coding. For this technique, a block transform (such as discrete cosine transform) is applied to a vector (or a subimage) of sampled values. The transformed coefficients are then scalar quantized.

Daniel S. Hirschberg

Digitized Audio and Video Signals

The information content of speech, music, and television signals can be preserved by periodically sampling at a rate equal to twice the highest frequency to be preserved. This is referred to as Nyquist sampling. However, speech, music, and television signals are highly redundant, and use of simple Nyquist sampling to code them is inefficient. Reduction of redundancy and application of more efficient sampling results in compression of the information rate needed to represent the signal without serious impairment to the quality of the remade source signal at a receiver. For speech signals, redundancy evident in pitch periodicity and in the format (energy-peaks) structure of the signal's spectrum along with aural masking of quantizing noise are used to compress the information rate. In music, which has much wider bandwidth than speech and far less redundancy, time-domain and frequency-domain masking are principally used to achieve compression. For television, redundancy evident in the horizontal and vertical correlation of the pixels of individual frames and in the frame-to-frame correlation of a moving picture, combined with visual masking that obscures quantizing noise resulting from the coding at low numbers of bits per sample, are used to achieve compression. *See* SPEECH; TELEVISION.

Compression techniques may be classified into two types: waveform coders and parametric coders. Waveform coders replicate a facsimile of a source-signal waveform at the receiver with a level of distortion that is judged acceptable. Parametric coders use a synthesizer at the receiver that is controlled by signal parameters extracted at the transmitter to remake the signal. The latter may achieve greater compression because of the information content added by the synthesizer model at the receiver.

Waveform compression. Waveform compression methods include adaptive differential pulse-code modulation (ADPCM) for speech and music signals, audio masking for music, and differential encoding

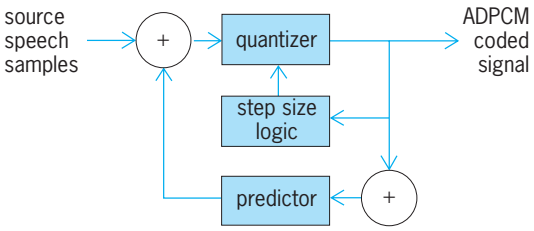

Fig. 1. Adaptive differential pulse-code modulation (ADPCM) encoder.

and sub-Nyquist sampling of television signals.

ADPCM. For coding of speech signals, this method (**Fig. 1**) incorporates predictors at the transmitter and receiver having response functions representative of the short- and long-term correlation inherent in the signal. The difference between the predicted signal and the input signal is coded by a signal quantizer that adapts to difference signal amplitude. The coded difference values are transmitted to the receiver and used to reconstruct the speech signal. The quality of the ADPCM-coded speech signal recovered at the receiver at 32 kilobits per second (kbps) is equivalent to that of pulse-code modulation at 64 kbps, and is suitable for public switched telephone service. ADPCM at 16 kbps has a lesser yet acceptable quality which is suitable for mobile telephone service. *See* PULSE MODULATION.

Coding for a bandwidth up to 7 kHz for speech and music is provided by an ADPCM coder that splits a 16-kHz-sampled, 14-bit, linearly quantized source signal into two 8-kHz critically sampled components, one for the low subband and the other for the high subband. Coding of the low subband uses 6 bits per sample yielding 48 kbps, and the high subband 2 bits per sample yielding 16 kbps, for a total of 64 kbps.

CD-quality coding. Good to excellent compact-disk (CD) quality for the reproduction of 14-, 20-, and 22-kHz bandwidth music signals has been achieved at coding rate rates of 2 bits per sample. The coding method is based on audio (music) signal masking of the quantizing noise. The input signal is coded by using a 16-bit-per-sample linear quantizer at 44.1 kHz for a 20-kHz-bandwidth input signal. The signal is windowed and converted to 28 equally spaced subbands, each having 750-Hz width, by a polyphase processor. The subband signals, sampled at 1.5 kHz, are quantized with a number of bits sufficient to keep the quantizing noise masked in each subband. A masking level in each subband is calculated by using a fast-Fourier-transform spectrum analyzer operating in parallel. The coded signal, comprising the sample bits and identifying the bit quantizer used for each subband, is sent to the receiver where an inverse polyphase processor remakes the digital audio signal. *See* COMPACT DISK; FOURIER SERIES AND TRANSFORMS; SPECTRUM ANALYZER.

Differential encoding. Television pixels are strongly correlated from line to line and from frame to frame. By coding these differences, the bit rate needed to transmit television can be reduced in excess of 4:1 depending on source material. At the receiver the differences progressively update a stored version of the frame that is displayed after each update. Motion prediction used in the difference coding can double the compression ratio. Differential coding can also be used with orthogonal transform coding as discussed below.

Sub-Nyquist sampling. This refers to sampling at a rate lower than Nyquist to preserve signal content without aliasing distortion. It can be applied to a television signal by sampling it at a rate less than Nyquist and at an odd multiple of the frame rate. This places the aliasing components into periodically spaced voids in the television spectrum where they can be removed by a comb filter at the receiver.

Parametric encoders. These include vocoders for speech signals and encoders using orthogonal transform techniques for television.

Vocoders. A vocoder analyzes the source speech signal at the transmitter to extract parameters suitable for control of a synthesizer that remakes speech at the receiver. A typical speech analyzer-synthesizer is based on a linear predictive coder (LPC) model of the human vocal cavity (**Fig. 2**). The synthesizer consists of an excitation source that presents a combination of periodic pitch pulses and noise to a dynamic vocal cavity model. The pitch pulses represent the excitation presented to the human vocal cavity by the vibration of the larynx, and the noise represents the excitation created as air flows through constrictions in the vocal cavity. The synthesizer has a transmission response which, under control of the parameters extracted at the transmitter, tracks the size and shape of the vocal cavity in the process of speech production. Compression occurs because these

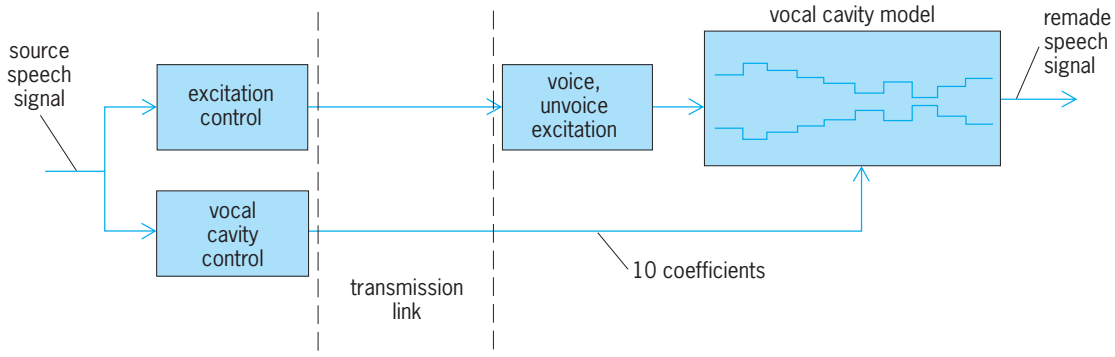

Fig. 2. Linear predictive coded (LPC) vocoder.

parameters vary at low rates compared to the high frequencies that appear in the source speech. Typical sampling rates for the parameters are 40–50 Hz. The number of bits per sample required to specify the parameters and the excitation is between 96 and 192, yielding transmission bit rates of 4800–9600 bits per second (bps) at 50-Hz sampling rate, compared with 64,000 bps typical of Nyquist sampling of telephone-quality voice signals.

The LPC synthesizer is equivalent to a cascade of tube sections, each section of fixed length and varying cross-sectional area, that approximates the size and shape of the vocal cavity. The remade speech of the LPC using 10 or 12 tube sections and excited by a multiband excitation (MBE) method is very natural at bit rates of 4800 and 9600 bps, but it still bears a trace of quality impairment. Multiband excitation dynamically assigns pitch excitation to some bands and noise excitation to other bands to excite the synthesizer. *See* VOICE RESPONSE.

Orthogonal transform techniques. Orthogonal transforms, such as discrete Fourier, Hadamard, and Haar, are used to convert blocks of $N \times N$ pixels into the coefficients of a set of N^2 orthogonal functions which are summed at a receiver to reconstruct a television picture. A two-dimensional discrete cosine transform (DCT), which is a Fourier transform, is widely used to code frame-to-frame differences for teleconferencing and broadcast-quality television. Typically, an array of 16×16 pixels, each pixel coded by 8 bits, is transformed into 256 DCT coefficients. In the reconstruction of the signal, some DCT coefficients must be coded with 10–12 bits, others with a lesser number, and some can be dropped altogether. On the average, 1 or 2 bits per coefficient are required. This represents a compression of 4:1 to 8:1, depending on the image content and the reproduction quality objective. When combined with frame-to-frame difference and motion prediction coding, overall compression ratios of 40:1 and greater result. Such techniques can also be used to achieve significant reductions in the transmission bit rate needed for high-definition television (HDTV). *See* INTEGRAL TRANSFORM.

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Data mining

The development of computational algorithms for the identification or extraction of structure from data. This is done in order to help reduce, model, understand, or analyze the data. Tasks supported by data mining include prediction, segmentation, dependency modeling, summarization, and change and deviation detection. Database systems have brought digital data capture and storage to the mainstream of data processing, leading to the creation of large data warehouses. These are databases whose primary purpose is to gain access to data for analysis and decision support. Traditional manual data analysis and exploration requires highly trained data analysts and is ineffective for high dimensionality (large numbers of variables) and massive data sets. *See* DATABASE MANAGEMENT SYSTEM.

A data set can be viewed abstractly as a set of records, each consisting of values for a set of dimensions (variables). While data records may exist physically in a database system in a schema that spans many tables, the logical view is of concern here. Databases with many dimensions pose fundamental problems that transcend query execution and optimization. A fundamental problem is query formulation: How is it possible to provide data access when a user cannot specify the target set exactly, as is required by a conventional database query language such as SQL (Structured Query Language)? Decision support queries are difficult to state. For example, which records are likely to represent fraud in credit card, banking, or telecommunications transactions? Which records are most similar to records in table A but dissimilar to those in table B? How many clusters (segments) are in a database and how are they characterized? Data mining techniques allow for computer-driven exploration of the data, hence admitting a more abstract model of interaction than SQL permits.

Data mining techniques are fundamentally data reduction and visualization techniques. As the number of dimensions grows, the number of possible combinations of choices for dimensionality reduction explodes. For an analyst exploring models, it is infeasible to go through the various ways of projecting the dimensions or selecting the right subsamples (reduction along columns and rows). Data mining is based on machine-based exploration of many of the possibilities before a selected reduced set is presented to the analyst for feedback.

KDD process. Data mining is a single step in the larger knowledge discovery in databases (KDD) process of identifying valid, novel, potentially useful, and ultimately understandable structure in data. Here, structure refers to either patterns or models. A pattern is an expression representing a parsimonious description of a subset of the data. A model is a representation of the source generating the data. The term "process" implies that knowledge discovery in databases comprises many steps. The step of data mining consists of the enumeration or identification of structures over the data. To be deemed knowledge, the derived structure must pass certain criteria such as utility (for example, gain, the amount of money saved by improved predictions) and validity (for example, estimated prediction accuracy). More difficult criteria to quantify include novelty and understandability. Sometimes various measures are combined as a single interestingness measure.

Density estimation and model selection. The fundamental problem of data mining is density estimation. If a density model of the joint probability density of the source that generated the data is available, then all questions about the data can be answered. Estimating the density, while a useful unifying view, is a difficult problem, and hence more specific tasks are often solved instead.

The other basic question is: Given a fixed set of data D, and many possible models for it, which model is to be selected? Formally, it is desirable to choose the model M for which the probability $\Pr(M|D)$ is maximized. Rewriting this probability using Bayes rule yields the equivalent expression given in

$$\Pr(M|D) = \frac{\Pr(D|M)\ \Pr(M)}{\Pr(D)}$$

which is known as the maximum a posteriori probability (MAP). The denominator of this expression is constant for the specified set of data, and $\Pr(M)$ is known as the prior probability of a model before the data are examined. The third term, $\Pr(D|M)$, is a measure of degree of the fit of the data to the model. When the prior probability is ignored, then maximum likelihood estimation is being performed, whereby the model is chosen that best fits the data. This procedure often leads to a problem known as overfitting the data with unjustifiably complex models. Maximum a posteriori probability hence gives a trade-off between model likelihood and fit to data. In statistics this approach appears under various names, including bias-variance trade-off. Taking logarithms of the expression in the above equation yields an equivalent form known as the minimum message (description) length, which calls for choosing the model that gives the minimal encoding of the data. The cost of minimally encoding an object M is at least $-\log[\Pr(M)]$ bits. Hence, minimizing $\log[\Pr(D|M)] + \log[\Pr(M)]$ prescribes choosing the model that minimizes the number of bits required to specify the model and the data, given the model. *See* BAYESIAN STATISTICS; PROBABILITY; STATISTICS.

Client-server decomposition. Algorithms for data mining build their models based on sufficient statistics or sufficient statistics in the data. Deployment of scalable data mining systems that can operate on large databases requires decomposing the operation into server-based operations that collect sufficient statistics (counts and data-specified parameters on models) and client-based operations that consume the sufficient statistics to produce models. This breakdown (see **illus.**) primarily avoids the unnecessary and expensive movement of large data sets. Unfortunately, SQL does not provide required sufficient statistics efficiently; hence, the server layer performs specialized data reduction operations. Traditional statistical or pattern recognition tools for data analysis fail on large databases since they assume that data fit in memory and hence do not scale. *See* CLIENT-SERVER SYSTEM.

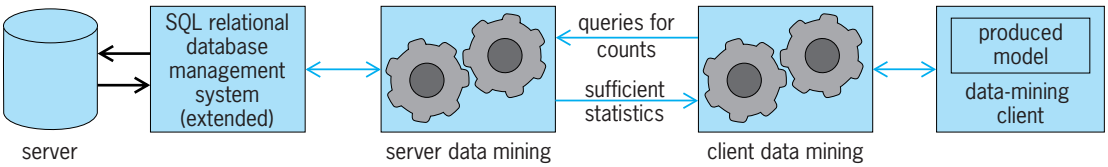
Methods. Data-mining methods include predictive modeling, segmentation, data summarization, dependency modeling, and change and deviation detection.

Predictive modeling. This method predicts some field or fields in a database based on other fields. Regression refers to prediction of numeric (continuous) variables. Classification refers to prediction of categorical fields. The problem is to determine the most likely value of the variable being predicted, given the other fields (inputs), the training data (in which the target variable to be predicted is given for each record), and assumptions representing prior knowledge.

Segmentation. This method, also known as clustering, separates the data items into subsets that are similar to each other. A two-stage search is employed: an outer loop over possible numbers of clusters, and an inner loop to fit the best possible clustering for a given number of clusters.

Data summarization. Extracting compact descriptions of subsets of the data can be done by taking horizontal (cases) or vertical (fields) slices of the data. Taking horizontal slices provides summaries of subsets (for example, sufficient statistics, or logical conditions that hold for subsets); taking vertical slices finds relations between fields as opposed to predicting a specified field (classification) or grouping cases (clustering). One common method is called association rules. Associations are rules that state that certain combinations of values occur with other combinations of values with a certain frequency and associated confidence. A common application of this is market basket analysis, where the aim is to summarize which products are bought with what other products. The number of possible rules increases exponentially with the number of products, but due to data sparseness, only a few such rules satisfy given support and confidence thresholds. Scalable algorithms find all such rules in linear time (that is, in a time proportional to the size of the data set) for reasonable threshold settings. *See* ALGORITHM.

Dependency modeling. Insight into data is often gained by deriving causal structure: either probabilistic (statements about the probability distribution



Client-server decomposition for data mining.

governing the data) or deterministic (functional dependencies between fields). In addition to density estimation, there are methods for explicit causal modeling (for example, belief network representations of density functions).

Change and deviation detection. These methods account for sequence information: time-series or some other ordering (for example, protein sequencing in genome mapping). Ordering of observations is important and must be accounted for. Scalable methods for finding frequent sequences in databases are exponential in complexity in worst-case scenarios. However, they appear to execute efficiently because of the sparseness of data in real-world transactional databases.

Usama Fayyad

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Data reduction

The techniques used to transform massive datasets into a more suitable, or smaller, form for data analysis, while preserving the intrinsic characteristics of the data and minimizing the loss of accuracy. Data reduction is used for scientific computation, statistical analysis, control of industrial processes, and business applications, as well as in data mining. Many data reduction techniques focus on obtaining an approximate representation of the data, others on reducing the size of the original data. *See* DATA MINING; DATABASE MANAGEMENT SYSTEM.

Data reduction includes both parametric and nonparametric techniques. Parametric techniques assume a model for the data and attempt to estimate the model parameters that produce a best fit of the data, while nonparametric techniques represent, or categorize, the data without making any assumptions on the data model. *See* STATISTICS.

Dimensionality reduction techniques. Dimensionality reduction techniques concentrate on reducing the number of dimensions (attributes) of the data, while preserving the original information as much as possible. One class of methods for dimension reduction seeks to project the data into a smaller subspace than the original data space. The most widely used method of this class is principal component analysis (PCA). Given a set of data points or a vector in a d-dimensional data space, PCA finds a set of orthonormal vectors that point in the direction of the variability of the data known as principal components. A subset of the principal components that represents the largest variability of the data constitutes a base in a subspace of lower dimension. The projection of a point into that lower-dimensional subspace is an approximation of the point with a minimum loss of accuracy. There are several methods to compute the principal components; for instance, single-value decomposition is one of the most popular methods.

Feature selection is another class of methods that achieves dimensionality reduction by attempting to determine the relevant attributes of the data without applying any transformation. These methods use a heuristic approach to determine a minimum set of relevant attributes in the data according to certain evaluation criterion. Feature selection techniques are classified as supervised whenever class information is available for the data or unsupervised otherwise. Some evaluation criteria for unsupervised feature selection include maximum information compression index and correlation between attributes. Dependency measures and mutual information are examples of evaluation criteria used in supervised feature selection. One simple approach used in supervised feature selection is variable ranking. In this approach, a score is assigned to each attribute to rank the relevance of the attribute with respect to the class distribution of the data. Those attributes that show the highest relevance are selected. Two other main strategies used by both supervised and unsupervised feature selection methods are stepwise forward selection and stepwise backward selection. Stepwise forward selection methods start with an empty set of attributes and build the set of relevant attributes incrementally. At each step, the methods add the best attribute according to the evaluation criterion. Conversely, stepwise backward selection methods begin with the set of all attributes, and at each step remove the worst attribute according to the evaluation criterion. A hybrid method combines both strategies by adding the best of the remaining attributes and removing the worst of the current selected relevant attributes.

Numerosity reduction. There are two main approaches for numerosity reduction; one focuses on estimating a representation of the data by fitting the data to some model, while the other concentrates on finding a smaller representation of the data. Some methods for numerosity reduction, such as regression, are parametric, while others, such as sampling, clustering, and histograms, are nonparametric.

Regression models. Parametric models can be used to generate an appropriate approximation of the data. In this context, data reduction is achieved by storing the parameters of the model that best fit the data. Among several parametric techniques, regression models represent the data as a function of a set of independent variables. The simplest regression model is known as multiple linear regression. In this model, a dependent target variable Y is modeled as a linear function of several independent variables $X_1, X_2, \ldots, X_n$ as $Y = \beta_0 + \beta_1 X_1 + \beta_2 X_2 + \cdots + \beta_n X_n$. The parameters $\beta_0, \beta_1, \ldots, \beta_n$ are estimated using a least-squares fitting procedure. Also, regression models can use nonlinear functions of the dependent variables. After a proper redefinition of the variables, several nonlinear regression models can be transformed into a linear regression model. However,

there are nonlinear regression models that cannot be transformed into linear models. Although, the least-squares estimate of the parameters of some of these nonlinear regression models can be computed, many of them are intractable.

Histograms. Histograms are an effective nonparametric technique to represent an approximation of the data distribution by partitioning the data into "buckets." A bucket is a multidimensional cube, with each dimension consisting of a continuous range of values that are grouped (binned) at equal intervals. Typically, the frequencies of the data values contained in a bucket are considered to be the same and are computed by averaging the frequency of the binned data values in the bucket. The buckets are determined by a partitioning rule. Several types of histograms have been proposed according to different partitioning rules such as equiwidth, equidepth, and v-optimal. Equidepth histograms partition the data into a predefined number of buckets, such that all the buckets have the same width. An equidepth histogram splits the data into a number of buckets in such a way that the frequencies of the data values are the same in each bucket, while a v-optimal histogram partitions the data into a fixed number of buckets to minimize the variance of frequencies.

Sampling. Sampling is another nonparametric technique that reduces the size of the original data by selecting a representative sample of the data. There are several sampling techniques: simple random sample without replacement (SRWOR), simple random sample with replacement (SRWR), stratified sampling, and cluster sampling. SRWOR draws a sample from the data, where each data point has an equal chance of being selected but can be selected only once. SRWR is similar to SRWOR except that a selected data point is placed back before the next data point in the sample is drawn; therefore, a data point can be selected more than one time. Stratified sampling partitions the data into mutually disjoint groups known as "strata." A sample of each strata is extracted using simple random sampling. Typically, the size of the sample for each strata should be proportional to the size of the strata. In cluster sampling, the data are grouped into disjoint groups or "clusters." Some of the clusters are selected using simple random sampling; the resulting sample includes all the data points belonging to the selected clusters.

Clustering. Clustering is a nonparametric technique that attempts to identify similar groups in the data according to certain criterion. In the context of data reduction, a representative of each cluster is selected to represent the data points in the cluster. Two types of clustering methods are hierarchical clustering, which can be agglomerative or divisive, and partitioning clustering. Agglomerative clustering methods build a hierarchy in a bottom-up fashion. They begin by considering each data point as a cluster. At each step, the methods select two clusters to merge according to the evaluation criterion. In contrast, divisive clustering methods start with one cluster that contains all the data points. At each step, it splits a selected cluster in two clusters. The process continues until no further splitting is possible. Partitioning clustering methods try to find the best partition of the data in a predefined number of clusters.

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Data structure

A means of storing a collection of data. Computer science is in part the study of methods for effectively using a computer to solve problems, or in other words, the entire process of determining exactly the problem to be solved. This process entails (1) gaining an understanding of the problem; (2) translating vague descriptions, goals, and contradictory requests, and often unstated desires, into a precisely formulated conceptual solution; and (3) implementing the solution with a computer program. This solution typically consists of two parts: algorithms and data structures.

Relation to algorithms. An algorithm is a concise specification of a method for solving a problem. A data structure can be viewed as consisting of a set of algorithms for performing operations on the data it stores. Thus algorithms are part of what constitutes a data structure. In constructing a solution to a problem, a data structure must be chosen that allows the data to be operated upon easily in the manner required by the algorithm. Because there are different ways to organize data and devise algorithms, it is important to develop criteria for choosing among them. Thus attention is also given to the analysis of the behavior of algorithms under various conditions.

Data may be arranged and managed at many levels, and the variability in algorithm design generally arises in the manner in which the data for the program are stored, that is (1) how data are arranged in relation to each other, (2) which data are calculated as needed, (3) which data are kept in memory, and (4) which data are kept in files, and the arrangement of the files. An algorithm may need to put new data into an existing collection of data, remove data from a collection, or query a collection of data for a specific purpose. *See* ALGORITHM.

Abstract data types. Each data structure can be developed around the concept of an abstract data type that defines both data organization and data handling operations. Data abstraction is a tool that allows each data structure to be developed in relative isolation from the rest of the solution. The study of data structure is organized around a collection of abstract data types that includes lists, trees, sets, graphs, and dictionaries. *See* ABSTRACT DATA TYPE.

Primitive and nonprimitive structures. Computers must have data before they can perform operations on them. Ultimately, data must be processed and organized in a manner that facilitates processing.

Data can be structured at the most primitive level, where they are directly operated upon by machine-level instructions. At this level, data may be character or numeric, and numeric data may consist of integers or real numbers.

Nonprimitive data structures can be classified as arrays, lists, and files. An array is an ordered set which contains a fixed number of objects. No deletions or insertions are performed on arrays. At best, elements may be changed. A list, by contrast, is an ordered set consisting of a variable number of elements to which insertions and deletions can be made, and on which other operations can be performed. When a list displays the relationship of adjacency between elements, it is said to be linear; otherwise it is said to be nonlinear. A file is typically a large list that is stored in the external memory of a computer. Additionally, a file may be used as a repository for list items (records) that are accessed infrequently.

Internal and host-language structures. In general, applications use representations of data based on internal data types defined by the computer architecture. For example, "char" is a single byte that is capable of holding or storing one character in the local character set; "int" is an integer, typically reflecting the natural size of integers on the host machine, 16 or 32 bits. Other applications work with host-language data types which are predefined by the language in which they are written. *See* COMPUTER PROGRAMMING; PROGRAMMING LANGUAGES.

Linear data structures. An array is a means of allocating space in a computer memory for data. Array allocation is static because once it is done it never changes. Static memory allocation works well when the number of data items to be operated upon, and thus the amount of physical storage that will be required for the task, is known. Dynamic allocation allows the number of data items (elements) to grow. Memory is allocated as needed, by request from the application, provided it is still available. When memory or storage is requested, the appropriate size (block) and type required by the data items will be set aside.

An array may be envisioned geometrically as a number of adjacent cells of equal size. **Figure 1** shows a linear array of integers with the name thisarray. Each element of the array has 32 bits or 2 bytes of data and the first element of the array is referred to as thisarray[1], the second element as thisarray[2], and so on. The *k*th element is thisarray[k]. In this example, the location of the first element is loc, the second element is at loc + 2, and so on. Thus the second element is 2 bytes removed from the first.

In the most general form of a linear list, deletion and insertion of an element can be made at any position in the list. A subclass of lists permits insertion or deletion to occur only at one end (**Fig. 2**). The last information in such a list is processed first, that is, on a last-in first-out (LIFO) or a first-come last-served

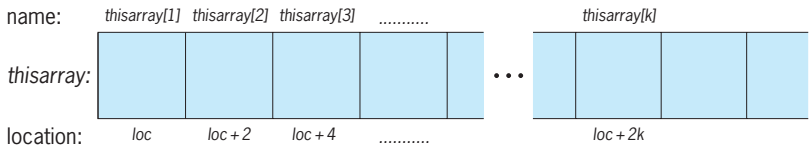

Fig. 1. Linear array of integers.

(FCLS) basis. A linear list belonging to this class is called a stack. The insertion operation is known as push, and the deletion operation is known as pop. A stack can be envisioned as a pile of trays. A tray is removed only from the top of the stack (pop) and returned only to the top of the stack (push). The most and least accessible elements are the top and bottom of the stack respectively, and the last tray in is the first taken out.

Another subclass of lists permits insertion at one end and deletion at the other (**Fig. 3**). The information in such a list is processed in the same order it was received, that is, on a first-in first-out (FIFO) or first-come first-served (FCFS) basis. This type of list is referred to as a queue. Each deletion removes the oldest remaining item in the list or structure. A line of patrons at a ticket window serves to illustrate a queue. Entry is made at the rear, exit at the front, and the person at the head of the list is served first.

Variations of this structure exist. A double-ended queue deque is a linear list in which insertions and deletions are permitted from either end of the structure. There are also two variations of the deque: the input-restricted deque, which permits insertions at only one end, and the output-restricted deque, which permits deletions from only one end.

These structures can be described by using the adjacent-allocation method of storage, which works well in situations when the storage requirement of the application is known. There are applications, however, for which the storage requirements are

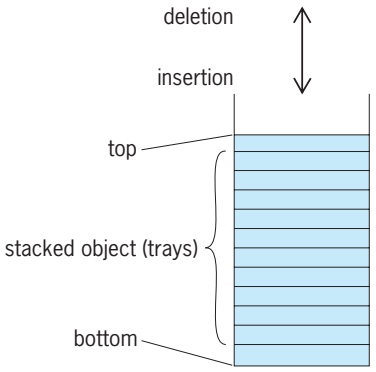

Fig. 2. A stack, a linear list in which elements are inserted or deleted only at one end.

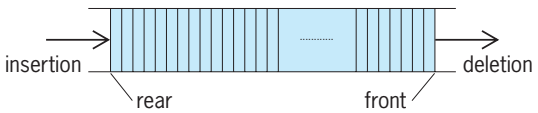

Fig. 3. A queue, a linear list that permits insertion at one end and deletion at the other.

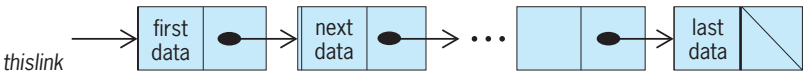

Fig. 4. Singly linked list or one-way chain.

unpredictable and cannot be determined at the time that the program is written; and those for which extensive data manipulations or operations such as insertions and deletions of data are required. In these cases the linked-allocation method of storage can result in the efficient use of both computer storage and time.

Linked allocation. The use of links to refer to elements of a linearly ordered data structure means that elements that are logically adjacent because of the linear ordering may not necessarily be adjacent in computer memory. These types of allocation are referred to as linked allocation. Each element in the list is represented by a node. In a singly linked linear list or a one-way chain (**Fig. 4**), each node of the list references the subsequent node, and thus, starting with the first node in the list, other nodes can be processed by repeatedly traversing the link or reference to the next node. Below thislink in Fig. 4 is a pointer which indicates the first node of the list. This first node contains data for the first element. The purpose of its link is to specify the next node in the linear ordering. The last node indicates the end of the list, and therefore its link address is NULL, which is indicated by a slash.

It is possible for a list to have no nodes at all. Such a list is referred to as an empty list, and the value of thislink in the example would be NULL.

Nonlinear data structures. Nonlinear data structures are capable of representing more complex relationships than that of physical adjacency. One important nonlinear data structure is a tree. A tree is a type of graph.

Directed graphs. A directed graph **G** can be defined as a pair **G** = (**V**,**E**), where **V** is a finite set of points called the nodes or the vertices of the graph, and **E** is a set of ordered pairs of **V**, called the edges of the graph. Thus a directed graph can be represented by a diagram (**Fig. 5**) in which the vertices are shown as points and the edges are shown as oriented line segments corresponding to the pairs of vertices. For example, if (u, v) is an edge, then in a directed graph the line segment uv is drawn with an arrowhead at the point v. Figure 5 shows the directed graph with the vertices given by Eq. (1) and the edges given by Eq. (2). *See* GRAPH THEORY.

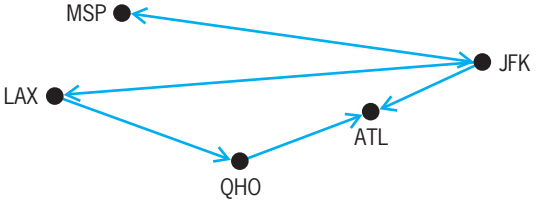

Fig. 5. Diagram representing a directed graph.

$$\mathbf{V} = \{\text{MSP, JFK, LAX, ATL, QHO}\} \quad (1)$$

$$\mathbf{E} = \{(\text{MSP,JFK}), (\text{JFK,MSP}), (\text{JFK,LAX}), (\text{JFK,ATL}), (\text{LAX,QHO}), (\text{QHO,ATL})\} \quad (2)$$

Paths. A path **P** is said to exist from a node u to another node v if there exist nodes $n_1, n_2, \ldots, n_k$ such that the pairs (u, n_1), (n_1, n_2), $(n_2, n_3), \ldots, (n_{k-1}, n_k)$, (n_k, v) are all edges of the directed graph. The collection of edges is called a path of the graph from node u to node v. Thus, in Fig. 5, there is a path **P**(LAX,ATL) from Los Angeles (LAX) to Atlanta (ATL) through Houston (QHO) given by the edges {(LAX,QHO), (QHO,ATL)}.

In the case of directed edges, the two possible edges between a pair of nodes which are opposite in direction are considered distinct. Thus like a two-way street, there is a connection between New York (JFK) and LAX and from LAX to JFK.

Trees. A directed graph is said to be a tree if (1) there exists a distinguished node R called the root of the tree that has no edge going into it, and (2) for every other node u in the graph, there exists exactly one path from the root R to the node u. In **Fig. 6**, for example, the path from R to u_6 is given by Eq. (3). Trees are useful for describing any structure

$$\mathbf{P}(R, u_6) = R \rightarrow u_1 \rightarrow u_2 \rightarrow u_6 = \{(R, u_1), (u_1, u_2), (u_2, u_6)\} \quad (3)$$

which involves hierarchy. Thus u_2, u_3, and u_4 are children or immediate successors of node u_1. Every node which is reachable from a node, n say, is called a successor of n. Also, u_5 is a successor of u_1, and u_1 is the immediate predecessor or parent of u_2, u_3, and u_4. Any node with no successor is called a leaf or terminal node. Thus nodes u_3, u_4, u_5, u_6, 3, and 4 are all leaves.

If the root R and the edges connecting it to the next level are deleted (removed) from the graph, a set of disjoint trees is obtained, rooted in the nodes which were the children of R. These trees are subtrees of the original configuration. A set of disjoint trees is a forest.

Although the nodes in a general tree may contain any number of pointers to other tree nodes, a large number of data structure applications involve binary trees, in which each node can have at most two pointers (successors) to other tree nodes. Various storage representations for trees are based on the linked allocation technique. In one linked representation of binary trees, there are two pointers from each parent node to its left and right subtrees, respectively (**Fig. 7**). An empty subtree is represented by a pointer value of NULL.

File structures. Not all information that is processed by a computer necessarily resides in immediately accessible memory because some programs and their data cannot fit into the main memory of the computer. Large volumes of data or records and archival data are commonly stored in external memory as entities called files. Any storage other than

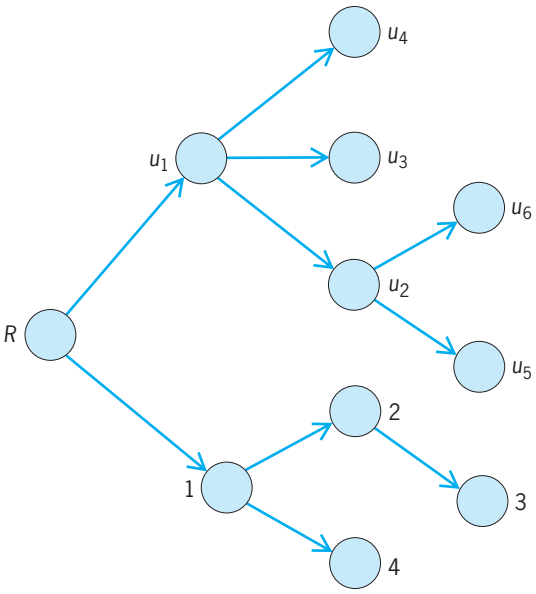

Fig. 6. Example of a tree.

main memory may be loosely defined as external storage. This includes tapes, disks, and so forth. A record item that uniquely identifies a record in a file is called a key. The most common types of file organizations are sequential, indexed sequential, and direct. Access of components of a file can be sequential or direct (random) access. *See* COMPUTER STORAGE TECHNOLOGY.

In a sequential file, records are stored one after the other on the storage device. In sequential access the values can be accessed only in the order in which they are stored in the file, while in random access the values can be accessed as the program desires. The operations performed on a sequential file may differ depending on the storage device used. A file on a magnetic tape, for example, can be opened for either input or output but not both at the same time.

Sequential processing is the accessing of records, one after another, in ascending order by a key or index item of the record. Sequential processing and serial processing are most effective when a high percentage of the records in a file must be processed; since every record in the file must be scanned, a relatively large number of transactions could be batched together for processing. Indexed sequential access combines a sequential key organization with indexed access. A disk can be considered to be divided up into a number of pages, which are contiguous blocks of information that can be accessed by the disk hardware, and each page generally holds many records. To vastly improve the speed of access, an index is maintained that specifies which keys belong to which pages on the disk.

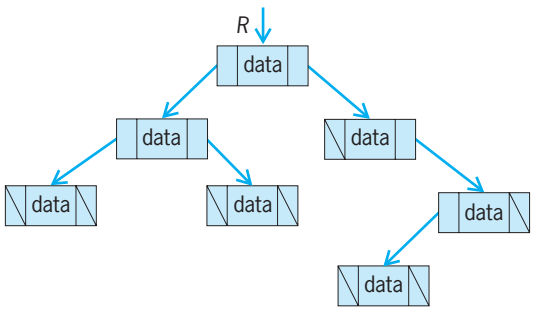

Fig. 7. Linked storage representation for a binary tree.

In a direct (random) access file, a transformation is required from the key of a record to the address of the storage location in which the record is resident. The transformation or mapping is referred to as a hashing technique and is made up of (1) a hash function, which defines a map from the key space to the address space, and (2) a collision-resolution technique to resolve conflicts when two or more record keys are mapped to the same location in a table.

Virtual memory. This is a system that provides an extension to main memory in a logical sense. Most modern computer systems provide a large virtual memory capability. In a virtual system, all currently active programs and data are allocated space or virtual addresses in virtual memory. The program and data may not in fact reside in main memory but in an external storage. References to virtual addresses are translated dynamically by the operating system into real addresses in main memory. This strategy relies on the fact that many programs have relatively small locality of reference; that is, each reference of a program to memory is likely to be in an area of memory that is relatively close to the other recently referenced areas, and thus transfer from external storage to main memory is infrequent. *See* DIGITAL COMPUTER.

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Database management system

The software that manages and controls access to data in a database. In a database management system (DBMS), the mechanisms for defining its structures for storing and retrieving date efficiently are known as data management.

A database is a collection of logically related data with some inherent meaning. In any modern enterprise where data and information are essential assets, databases play an important role. The power of databases comes from their conveniently and efficiently managing large amounts of data. For example, a telephone book can be considered a type of database, where the data concerning the names of

individuals and businesses are related to their corresponding telephone numbers and addresses.

Nowadays, a database generally refers to a computerized database with a collection of related data and software package or DBMS that manages and controls access to the data. A computerized database can be considered as an electronic "filing cabinet" for storing data/information and allowing users to manipulate that data/information on demand. Examples of computerized databases include airline reservation systems, banking systems, company payroll, and employee information systems.

The first commercial databases appeared in the late 1960s. These databases evolved from traditional file systems. Databases are also used in many different scientific or human endeavors such as the Human Genome Project, pharmaceutical research, anthropologic investigations, and nuclear medicine, to name some.

Database components. A computerized database has hardware, software, data, users, and procedural components.

Hardware. The hardware of a database is the actual computer system used for storing, accessing, and processing the data, as well as the different storage media on which the data are stored. In large organizations, the hardware includes a network with one or more servers and many client programs. Clients are programs that interact with the DBMS and may run on personal computers at the user's end. The server is generally the computer where the DBMS resides (**Fig. 1**). In general, this server has a more powerful processor than the client machines because it handles the data-retrieval operations and most of the actual data manipulation. *See* CLIENT-SERVER SYSTEM; COMPUTER SYSTEMS ARCHITECTURE; MICROPROCESSOR.

The secondary storage media where the data reside can accommodate large amounts of data and are nonvolatile; that is, any storage medium that, unlike main memory, can retain information in the absence of power. The most commonly used media are magnetic disks (hard disks) that can store large amounts of data at a lower cost than main memory. To improve their reliability, hard disks may be organized as redundant arrays of independent disks (RAID), where data are stored with some amount of redundancy across several disks. In the event of a disk failure, data can be recovered from the data on the other disks that form the system. Other media commonly used to store data include magnetic tapes, flash memory, optical storage, and DVDs. *See* COMPUTER STORAGE TECHNOLOGY; MAGNETIC RECORDING; OPTICAL RECORDING.

Software. The main software component of a database is its DBMS. This software package is the mediator or intermediate layer that allows users to retrieve and manipulate (create, delete, update, and modify) the data. In general, all operations on the data are performed through the capabilities of the DBMS. The DBMS also controls simultaneous access to the data and maintains its consistency or integrity.

The DBMS runs on a computer under a particular operating system. A DBMS and its clients can reside in a computer where it can be accessed by a single user (single-user systems) or by several users simultaneously (multiuser systems). In a network, the DBMS allows for data-handling programs to reside on the server and clients programs on each desktop. A DBMS also makes use of utilities, such as report writers, application development tools, and other design aids. A database is generally referred to by the name of its DBMS, for example, an Oracle® database. *See* OPERATING SYSTEM; SOFTWARE.

Data. The data component is the heart of any DBMS. There are two kinds of data. First, and most obvious, is the set of related information needed by the organization. The second kind is information about the database itself, known as metadata. This information is usually kept in the part of the DBMS called the data dictionary, or catalog. For every piece

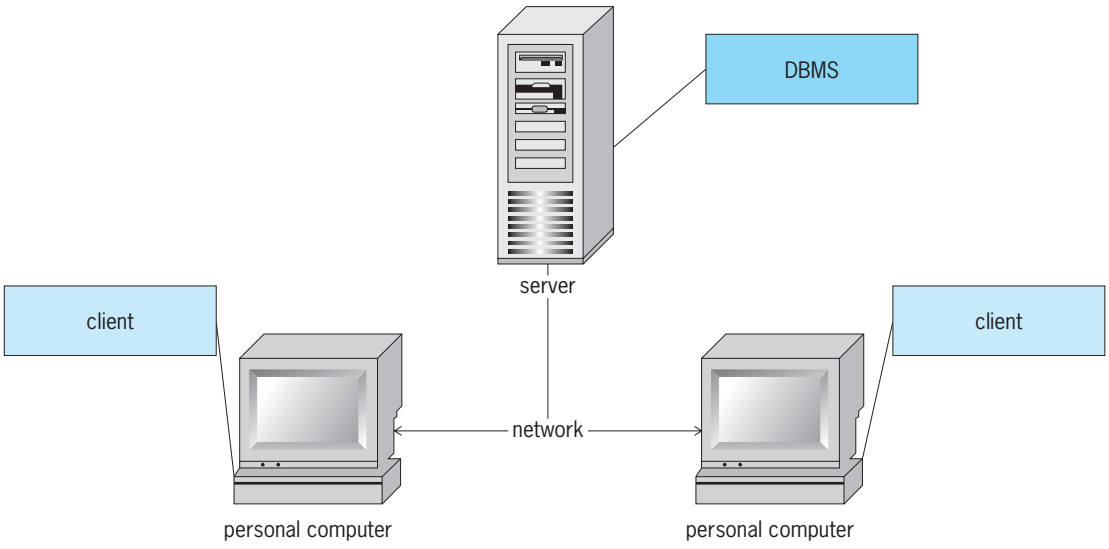

Fig. 1. Components of a database.

of data, the data dictionary holds information such as name, type, size, format, and any possible constraint (the conditions that the data must meet or satisfy). The data dictionary also contains information about all the users and the access type and actions that they are allowed to perform (privileges) in the database. *See* METADATA.

One of the major features offered by a DBMS is that the actual data are separated from the programs that use that data. The set of facts represented in a database is called the universe of discourse (UOD). The UOD is the representation of the "world" that is relevant to the database users. For this reason, databases are designed and built for a particular audience and for a specific purpose.

In a database, the things that are of interest to its users are called entities. The characteristics that describe or qualify an entity are called its attributes. For example, in a database that holds information about students at a particular university, the basic entity is the student. Among the attributes that describe a student entity are the student's first and last names, date of birth, student ID, major, department, and grade point average. There may be some other pieces of information that can be gathered about a student such as eye color, race, and height, but this information may not be relevant for this particular database.

The entities and attributes that describe them are part of the UOD and are stored in the data dictionary. The set of possible values that an attribute can have is called the domain of the attribute. For example, the attribute describing a student's field of study at a particular university may have values such as medicine, architecture, electrical engineering, or computer science. Each attribute of an entity must have a domain.

The entire description of all objects in the database is called the schema. The description of this schema depends on the data model (organizational guidelines) used and is specified during the design phase of the database. The schema should reflect the requirement specification of the database.

The data present in the database at any particular moment in time are called an instance or snapshot. The instance changes whenever new data are inserted or when the existing data are modified or deleted, keeping in mind that the data stored in the database may be required to satisfy certain conditions or constraints. For example, a constraint may be that every student must have a last name, a first name, and middle initial; that is, these attributes cannot be "left blank." For students who do not have a middle name, this attribute may be "filled in" with a predetermined value such as the letter "X." Another example of data constraint may be that the grade point average cannot be negative or exceed a predetermined value, say, 4.00.

Users. Even though databases are used in many fields and for many purposes, there is no such a thing as a general database for everything. Databases are built for specific purposes and for an intended user. For example, it is impossible to use a database built to operate an airline reservation system to manage a payroll system.

In a database, users can access or retrieve data on demand using the applications or the interfaces provided by the DBMS. In general, the users of a database can be categorized as application programmers, end users, or database administrators. Application programmers interact with the database through programs (applications) written in programming languages such as C++, Visual Basic, and Java. Applications generally extend or customize the original functionalities of the DBMS. *See* PROGRAMMING LANGUAGES.

End users are the people at workstations that have access to the data through the facilities provided by the DBMS. They may respond to requests from within or outside an organization. End users can take advantage of some of the interfaces provided by the DBMS to retrieve or update the data stored in the database. Some of these interfaces are menu-driven or forms-driven. Although somewhat different in appearance, these types of interfaces allow the user to select one type of action from a set of choices or to specify certain criteria that need to be met by the data being retrieved. In general, an end user is required to provide some type of input information to complete the action that has been selected. In a database of telephone subscribers, for example, a user can retrieve the address of a particular customer by inputting the telephone number.

Users retrieve data from a database by posing questions through the DBMS. These questions are called queries. Interfaces that allow the end users to query the database in plain English or any other spoken language are called natural language interfaces. This type of interface is still in development due to the inherent difficulties associated with having a computer communicate directly with a person. A somewhat primitive application of this type of database can be found in the services provided by some cellular telephone companies. Instead of dialing a number, users of these services talk on the telephone and mention the name of the person they want to call. Voice recognition software translates this verbal request into a search for a name in a database. *See* SPEECH RECOGNITION.

Another database interface frequently used by cashiers or automatic teller machines (ATM) is the parametric interface. Parametric interfaces, often used for simple and repetitive operations, generally require that the user press buttons on a screen or on special keypads. Each button or key triggers the execution of a procedure designed to perform a particular function.

Procedural components. Database administrators are the users responsible for defining the internal structures to store the data, the policies for manipulating the information, and the security mechanisms to protect the data. Database administrators also serve as liaison between end users and application programmers. The database administrator can be a single individual for small databases or a group of two or more people for large corporations.

An integral part of any database management system is the set of procedures that control the behavior of the system, that is, the actual practices the users follow to obtain, enter, maintain, and retrieve the data. For example, in a payroll system, a procedure may determine how salaries are calculated for the different types of employees that may exist in the organization. Part of the job of the database administrator is to verify that the procedures are clearly defined and comply with the guidelines of the organization.

Size of a database. The size of a database is generally measured by the size of its data. Databases can range from a few megabytes (10^6 bytes) for personal databases to gigabytes (10^9 bytes) or even terabytes (10^{12} bytes) for large corporate databases. Petabyte (10^{15} bytes) and exabyte (10^{18} bytes) databases will be common in the future.

Access. A DBMS provides a programming interface and simultaneous access to the data by more than one user. Applications programmers or data administrators can use the programming interface to manipulate or maintain the database. The DBMS allows concurrent access to data while making sure that the users do not interfere with one another.

Characteristics of the data in a database. In any database, the data are both integrated and shared. Integrated means that the database can be thought of as the unification of several distinct data files with redundancy between these files kept to a minimum. For example, a company database might contain data about its employees in an EMPLOYEE data file and information about the employees' dependents in a DEPENDENT data file. The employees' data might include employee ID, last name, first name, department, and salary. Data about the employees' dependents might include employee ID, dependent ID, dependent last name, dependent first name, relationship to employee, dependent date of birth, and so on. Information about the dependents of a particular employee is obtained by combining information from these two files using the common piece of data that they share (employee ID). The employee ID (or any other unique piece of identification) is necessary in this case to associate an employee with dependents since they are kept in separate files.

Shared data means that different users can use the same data for different purposes. For example, the human resources department might use the date of birth of the dependents to send them birthday cards, whereas the accounting department might use the same piece of data for tax purposes.

Data models. In a database, data have certain physical organization on the storage media on which they are kept. The data also have a logical organization as seen at the user interface of the DBMS. A major purpose of a database is to provide users, through its DBMS, with an abstract representation of the data. That is, users can conceptually picture the data in a particular manner while the system implements and hides all details on how the data are physically stored. For this purpose, databases are based on a data model.

A data model is a set of guidelines for representing the logical organization of data in a database. In other words, a data model aids database users by helping them think of, visualize, and communicate information about the data. A data model consists of named logical units and relationships among the data. One of several data models can be used for a particular database. The main difference between these data models is the manner in which they represent certain relationships among the data. Most data models handle attribute relationships in a similar manner but handle relationships (associations) between the entities in different ways.

Data models can be categorized according to the concepts that they use to describe the database. High-level or conceptual data models provide concepts that are close to the way users perceive the data, whereas low-level or physical data models provide concepts that describe the details of how the data are stored in the computer. In relational databases, for instance, all the data are viewed by the users as tables, while the data can be stored in a variety of complex data structures that permit efficient retrieval.

In addition, there is a class of implementation data models that use concepts understood by the end users with the data organized similarly within the computer. Implementation models are the most widely used in traditional commercial databases. These data models include the widely used relational data model and some of the legacy (older) models such as network and hierarchical models. However, the complexity of database applications has increased. A new family of higher-implementation data models has been proposed to address issues such as new data types for storing images, more complex structures for user-defined objects, and the need to define nonstandard application-specific operations. Another reason for new data models is to ensure their integration with object-oriented programming languages. The need for additional data modeling features has been recognized by some of the relational DBMS vendors who now offer extended or object-relational DBMSs. *See* DATA STRUCTURE; OBJECT-ORIENTED PROGRAMMING.

Three-level architecture. To facilitate the differentiation of the conceptual views of the data from the physical implementation, the American National Standard Institute (ANSI) has defined a general framework known as the three-level architecture (**Fig. 2**). Each level of this architecture, described by its own schema, serves a particular purpose. The internal level or physical level is concerned with the way the data are physically stored and accessed. The external level or user logical level is concerned with the way the data are seen by individual users. The conceptual level or community view describes what data are stored in the database and the relationships among the data. This architecture is a framework for describing the data and it is not the actual data. The data are physically stored in the computer system.

The three-level architecture can also be used to explain the concept of data independence—the ability

to change the schema at any level of the database without having to change the schema at the next higher level. There are two types of data independence: logical and physical. Logical data independence allows changing the conceptual schema without having to change the external schemas or applications programs. The conceptual schema may change by adding or deleting constraints to the existing data. Physical data independence allows changing the internal schema without having to change the conceptual schema. Changes to the internal schema may be the result of the reorganization of some of the physical files for storing the data.

Most DBMSs do not make explicit separation or support the three-level architecture in its entirety. In general, those DBMSs that support user views may support external schemas through the same data model that describes the conceptual level. The process of transforming requests and results between levels is called mapping. All information about the schemas and mapping between them is stored in the data dictionary. Whenever the schema at a lower level is changed but not the one at the next level, the notion of data independence is achieved by only changing the mapping between these two levels.

Data languages. Database designers and administrators use data languages to specify conceptual and internal schemas, as well as any mapping. A data language is a set of operations that permit access to the data as organized by a data model. A complete programming language that exists only to support the data language is called a self-contained data language. For example, DBMSs organized according to the relational data model use the self-contained data language known as the Structured Query Language (SQL). English-like, self-contained data languages are sometimes called query languages. This type of language allows the user to query the database using commands that follow a simple but somewhat rigid syntax. SQL is the de facto industry standard query language.

The subset of commands of the data language that allows the definition of the conceptual schema is called a data definition language (DDL). Another subset of the language that allows the definition of internal schema is called the storage definition language (SDL). Within the data language, the set of commands that allow the retrieval, insertion, deletion, and modification of the data is known as the data manipulation language (DML). If DML commands are used within an application program written in a high-level programming language, such as Java or C++, this programming language is referred to as the host language and the DML is called a data sublanguage (DSL).

Database classification. DBMSs can be classified according to criteria such as single-user or multiuser systems, as well as on the basis of the data model that they support. In addition, DBMSs can be classified according to where the DBMS and data reside. When both reside on the same computer system, the database is said to be centralized. If the DBMS and the data are distributed over many sites connected

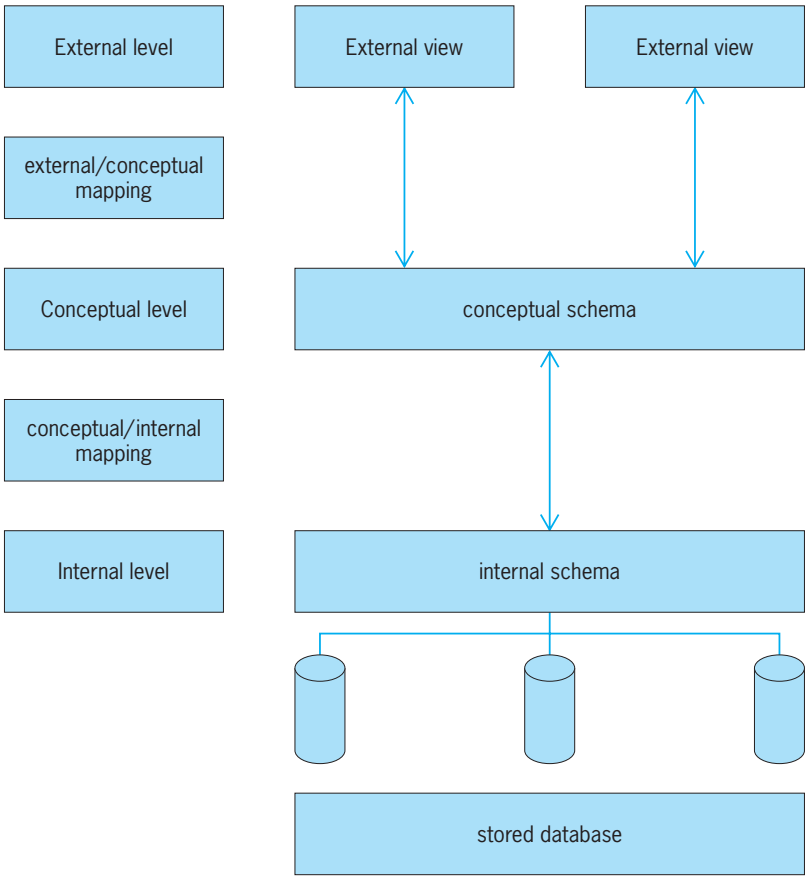

Fig. 2. Three-level architecture. (*Elmasri and Navathe*)

by a computer network, the database is said to be distributed. A distributed database is called homogeneous if all the DBMSs at the sites are the same; otherwise, the database is called heterogeneous. A federated database is a group of autonomous and independent databases that are connected through a network. A database can also be classified as special purpose if, for performance reasons, it is built for a specific application such as a medical registration database or telephone directory system. *See* DISTRIBUTED SYSTEMS (COMPUTERS); LOCAL-AREA NETWORKS; WIDE-AREA NETWORKS.

Client-server architecture. In client-server or two-tier architecture, client programs communicate with the DBMS across a network. The client program resides at the user's computers and the DBMS resides at the server computer. The client program can execute independently on its host computers. When a client program needs to communicate with the DBMS, it does so by using a special type of software that lets the client and the DBMS "talk" to each other. This special software has at least two components: one that runs on the client machine and another that runs on the server. These software components follow a predefined standard that defines an application program interface (API). An API consists of a set of library functions that implement many of the common operations that programmers require. Some of these operations include connecting to the database, making requests to the server, and obtaining results.

The industry standard API, based on the SQL language and developed by Microsoft, is a technology known as Open Database Connectivity (ODBC). The ODBC standard provides a common interface for accessing heterogeneous SQL databases. Through this interface, a single application can access data in different databases with a common set of codes. A specialized piece of software often used to facilitate this process is known as a driver. Drivers are provided by the different DBMS vendors and work exclusively with the DBMS for which they were created. A similar standard, called Java Database Connectivity (JDBC), has been created for accessing relational DBMS using Java applications. JDBC defines an API that supports basic SQL functionality and provides access to a large variety of relational DBMS products.

Many Web applications use a three-tier architecture where an additional layer of software is added between the clients and the server. This new layer, depending on the type of application, is called the application server or Web server. The Web server (which runs on its own computer) is the depository of procedures and constraints that have been defined to access the data stored at the database server. The Web server accepts the requests from the client, processes the request, and sends appropriate commands to the DBMS. The Web server receives the results from the DBMS and passes them to the client for further processing before being presented to the user. The Web server can also be used to secure the database. It can do this by checking the validity of the user's requests before sending them to the server or by using encryption/decryption mechanisms to protect the data. *See* COMPUTER SECURITY; WORLD WIDE WEB.

Relational model. The relational database management system has become the most popular of all the DBMSs. It was initially proposed by E. F. Codd in the early 1970s. In this model, all data are logically structured as relations or two-dimensional tables. Each entity of the universe of discourse corresponds to a table; the attributes of this entity form the columns of the table (**Fig. 3**). Each row or "tuple" of the table represents an instance of the entity. The values under each column must belong to the domain of that attribute but, in some cases, these values may be absent or "null." The table name and the column names help in interpreting the meaning of the values of each row. The total number of columns of the table is called its degree. The total number of rows in the table is called its cardinality.

The relational model is based on the mathematical concept of relation and uses some of the elements of predicate logic and set theory. However, some of these definitions, when compared to their mathematical counterparts, have been relaxed to accommodate the logical information conveyed by the table. *See* LOGIC; SET THEORY.

Any relation must have a set of one or more attributes, called the primary key, that uniquely identify each row. A foreign key is a column in a table that matches the primary key column in some other table. Key values in any table must be unique and cannot be null. When representing a relation as a two-dimensional table, the attributes that comprise the key are generally underlined.

Keys, with foreign keys, can also be used to maintain the consistency of the data among the tuples of two relations. In Fig. 3, the EMP_DEPT (employee department) attribute of the EMPLOYEE table has been defined as a foreign key for this table. This EMP_DEPT attribute references the primary key (ID) of the DEPARTMENT table. What this implies is that for any employee inserted into the EMPLOYEE table, the values under the department column must be one of the values listed in the key attribute of the DEPARTMENT table. The foreign key concept allows the EMP_DEPT to be NULL. However, if this NULL value is ever changed it has to be to one of the values listed in the ID column of the DEPARTMENT table.

A basic set of table operations, known as relational algebra, allow the user to specify basic retrieval requests. A SELECT operation extracts a subset of table rows that satisfy a given condition. A PROJECTION operation extracts some table columns. A JOIN operation combines related rows from two or more tables (not necessarily different) into single rows. Other operations on tables are the INTERSECTION, UNION, and DIFFERENCE. These table operations behave just as their counterpart set operations.

If for security reasons there is a need to hide the identity of the tables that participate in a query, users can create virtual tables or views. A view is a relation that can be produced on demand but does not necessarily exist in the database. Technically, views are queries stored in the data dictionary that users can "run." Views are a basic security mechanism.

Emerging technologies. Relational DBMSs are the most popular of all DBMSs. Estimated yearly sales of RDBMs and related products are around U.S. $20 billion. In the foreseeable future, relational DBMS will continue being the dominant DBMS. However, there are some new advances in database technology. Among the most promising technologies are mobile, multimedia, and geographical information systems. Mobile databases comprise portable computing devices, which, coupled with wireless communication, allow users to access data from virtually anywhere and at any time. Multimedia databases store and process data, including text, graphics, animations, video, audio, or any combination of these.

Department

ID	Name	Location
10	Accounting	New York
40	Sales	Miami

Employee

EMP ID	EMP_NAME	EMP_MGR	TITLE	EMP_DEPT
1234	Green		President	40
4567	Gilmore	1234	Senior VP	40
1045	Rose	4567	Director	10
9876	Smith	1045	Accountant	10

Fig. 3. Example table for a relational database.

See MULTIMEDIA TECHNOLOGY.

Geographical information systems (GIS) collect, model, store, and analyze information describing the physical properties of the world. GIS manipulate spatial data (such as maps, images from satellites, roads, weather, and soil) and nonspatial (such as census results, sales or marketing results). *See* GEOGRAPHIC INFORMATION SYSTEMS. Ramon A. Mata-Toledo

Bibliography. C. J. Date, *An Introduction to Database Systems*, 8th ed., Addison-Wesley, 2003; R. Elmasri and S. Navathe, *Fundamentals of Database Systems*, 4th ed., Addison-Wesley, 2003.

Dataflow systems

An alternative to the sequential, or von Neumann–based, model of computing, offering an inherently parallel metaphor. The dataflow concept is characterized by an orientation toward producing and consuming data values rather than updating storage cells. Such systems are made up of dataflow machines which execute programs expressed as a partial ordering of operations. This partial ordering, called a dataflow graph, is created by high-level language compilers. Dataflow systems are capable of exploiting all of the parallelism inherent in an algorithm. The economies of scale of very large scale integration can be exploited quite naturally by using the dataflow approach. *See* ALGORITHM; INTEGRATED CIRCUITS.

Basic concepts. Dataflow systems employ an underlying execution model which differs substantially from the conventional sequential model; specifically, the dataflow model is concerned with naming and manipulating data values as opposed to naming and manipulating the containers for data values, that is, storage cells. This simple yet fundamental change in perspective has a profound impact on issues both at the level of the programming language and at the level of the computing machinery. Moreover, the shift of focus from where and when values should be stored to merely what the values should be has a character which is mathematically more direct and appealing. This value-oriented view is called definitional: each step in the computation defines some entity, such as a number or a data structure, based on other such definitions or on input values. The traditional view is called imperative: each step in the computation not only defines something to be computed but also constrains where the resulting value should be stored and when such storing should take place. Because value containers can be reused within an imperative framework, additional constraints may be imposed implicitly on the time ordering of these steps.

Within the definitional framework, the statement

$$A = B + C$$

simply means that, in the scope of the identifier A, its value is defined as the sum of the values of B and C. It indicates nothing about where this value should be kept, nor even that the value must be computed at all, in the most rigorous sense. References to A within its scope all denote the same value, regardless of the time-ordering of the actual operations required to produce and consume the value. Given this view, statements such as

$$A = A + 1$$

are obviously meaningless as definitions in that they have no time-invariant denotation. This property in a programming language is sometimes called single assignment and is a key element of the dataflow model.

Another aspect of the definitional framework is that all operators behave functionally; the output value they produce is entirely defined by the input values with no reliance on the time at which the operation is performed (state independence). Moreover, operators and their compositions are constrained to be free of side effects.

These properties, when present in a programming language, allow compilers to deduce the essential dependences between operations and to direct dataflow machines to execute instructions accordingly. When the compiler can prove that two operations have no dependence of one upon the other, a dataflow machine can execute them in parallel.

Dataflow graphs. In the traditional model of computing, compiled instructions are represented as a sequential list. In the dataflow model, instructions are not totally ordered in this way but must be expressed as a partial ordering. Conceptually, each instruction can be viewed as a node in a graph. Directed arcs connect any instruction to those subsequent instructions which depend upon its value. Executing a program as a dataflow graph can be viewed as a process of propagating data-carrying tokens along the arcs of the graph. Tokens consist of a value and a tag. The tag serves to route the token from the instruction which produced the value to the instruction which is to consume the value. Tags denote an instruction within a graph, and in some classes of dataflow machines further specify a unique instantiation name which allows multiple invocations of the instruction to coexist. It is the dataflow machine's responsibility to execute, at any given time, all and only those instructions in a graph for which a complete set of identically tagged input tokens exists (so-called enabled instructions). An enabled instruction represents the complete specification of some processing activity: input values and function. Executing an instruction means computing the function denoted by the instruction (for example, *ADD*) by using the values on the input tokens, and fabricating a result token containing the new value and a new tag. This new tag denotes the instruction, called the destination, which is to receive the result. To the extent that there are many such enabled instructions at any given time, the machine is free to execute them in any temporal order whatever or, most interestingly, to execute them all in parallel. This execution rule is in contrast with the traditional method based on program counters.

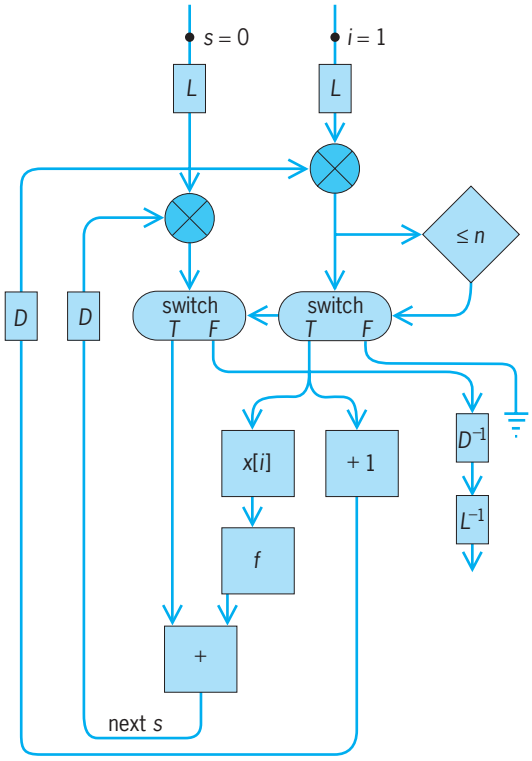

Fig. 1. Dataflow graph for compilation of the loop expression for $\sum_1^n f(x_i)$. (*After Arvind and R. A. Iannucci, Two Fundamental Issues in Multiprocessing: The Data Flow Solution, Computation Structures Group Memo 226-3, MIT Laboratory for Computer Science, 1985*)

Figure 1 is an example dataflow graph, compiled from a program fragment(shown below) in Id, a high-

```
{ s = 0
 in
 {for i from 1 to n do
  next s = s + f x[i]
  finally s }}
```

level language designed specifically for data-flow systems. The fragment applies the function f to successive elements of a one-dimensional array x and sums the results. The graph is somewhat stylized: the box marked f represents the subgraph necessary for computing function f. Instructions D, D^{-1}, L, and L^{-1} are tag-manipulation instructions which manage dynamic instantiation information on the tokens; they act as identity instructions for values. The *SWITCH* instruction routes a data token (top input) to either the T (true) or F (false) output, depending upon the value of a boolean token (right input). The remainder of the operators are arithmetic and relational instructions whose function should be self-evident. (The circled X represents the merging of two arcs and does not represent an instruction.)

Architectures. Clearly, a program could be designed to interpret dataflow graphs on a conventional architecture with the attendant inefficiencies that come with interpretation. Dataflow machines, on the other hand, compute by using encoded dataflow graphs as their machine language. Such a machine is shown in **Fig. 2**. As a token arrives at the waiting-matching unit, its tag is compared to those of the other tokens which are already waiting. Conceptually, this can be done with associative matching, although hash-based and direct-mapped techniques are used in practice. If no match is found, the token joins the waiting pool. If a match is found, both tokens are extracted and fed, along with their common tag, to the instruction fetch unit. Here, the tag is used to look up the correct instruction from the program memory. The values on the two tokens, along with the operation code from the instruction, are passed on to the arithmetic-logic unit (ALU), while the tag and the destination information are forwarded to the compute tag section. For most instructions, new tags are derived from old ones simply by substituting destination instruction information. Instructions such as D and L in Fig. 1 cause more sophisticated transformations on the instantiation-specific part of the tag, effecting the allocation and deallocation of instantiation names. The new value from the arithmetic-logic unit is combined with the new tag from the compute tag section, the result being a new token which is passed on to the waiting-matching unit. Parallelism can be exploited by distributing the dataflow graph across multiple processors. The routing network coordinates the task of shuttling tokens to the correct processors, based on their tags.

In this type of architecture, called dynamic dataflow, storage for the intermediate results of the computation, tokens in particular, is allocated at run time. Because of this, dynamic procedure invocation, including recursion, can be supported. For certain classes of computations wherein the intermediate storage issues can be worked out at compile time, a more restricted class of architecture, called static dataflow, is suitable. By preallocating storage and other resources, static machines avoid the need to carry instantiation information in the tags. *See* COMPUTER SYSTEMS ARCHITECTURE.

Suitability for parallel processing. The dataflow model of computation differs significantly from that of conventional systems; definitional semantics displace imperative semantics, and the data-driven notion of instruction scheduling displaces the notion of sequential control flow. It has been argued that the dataflow model is inherently parallel, whereas the conventional model is inherently sequential and that as a result dataflow is a more suitable basis for general-purpose parallel computing. A general-purpose parallel computer is, by its very nature, a collection of processors and memories distributed in space. If a single computation is to be mapped onto this set of processors, the fraction of the total computation (instructions and data) local to any one processor must necessarily decrease as the number of processors is increased. There is a direct relationship between this fact and the increase in latency as the degree of parallel processing increases. Latency is the time that elapses between making a request for a remote value, for example, and receiving the associated response. Because latency is

unavoidable, an important measure of any parallel processing architecture is its ability to tolerate long latencies.

Just as physically partitioning the machine (processors and memories) into n component parts gives rise to this fundamental issue of latency, so also decomposing the program into asynchronously communicating fragments gives rise to the fundamental issue of synchronization. This is not an issue in a conventional single-processor system, because there is a single locus of control. All instruction executions are implicitly synchronized through sequential instruction dispatching. Using such conventional machines in a parallel processor, however, invalidates this assumption. Synchronization may be provided, at some loss of generality, through a globally synchronous parallel processor architecture. A more universal approach necessitates the emulation, in software, of the synchronization means. Emulated synchronization is costly and must be used sparingly. The upshot for parallel processing is that, except for the most trivially parallelizable applications, seeking to minimize synchronization is tantamount to seeking to minimize the exploitation of parallelism. The most general approach, adopted under the dataflow paradigm, is to integrate synchronization features into the lowest levels of the individual processor in such a way that explicit synchronization takes place prior to the invocation of each instruction. This architectural step having been taken, the cost of synchronization is not a first-order programming concern. Providing efficient hardware means for synchronization is, however, a very significant hardware concern. *See* CONCURRENT PROCESSING.

Advantages and limitations. Dataflow systems offer an approach to scalable, general-purpose parallel processing which has been shown to address satisfactorily the fundamental issues of latency and synchronization. The approach relies on the data-driven dispatching of instructions and the manipulation of values, rather than value containers, to decouple the issue of temporal computing behavior (the time-interleaving of parallel activities) from the correctness of the computed result. Compared to conventional systems in which synchronization must be provided in software, dataflow machines provide synchronization which, for better or worse, is inherent in the scheduling of every instruction. This has a profound impact on the economics of exploiting high degrees of parallelism. Dataflow languages obviate the need for explicit parallelism notations in programs and the attendant concerns over correctness.

As in any parallel processing system, conventional or dataflow, the issues of dynamic resource management, such as the allocation and deallocation of storage and processors, must be considered. The exploitation of parallelism in a given application usually requires more resources than the sequential execution of the same application. Moreover, in the dataflow model, the problem of bounding or limiting the instantaneous resource requirements during execution is of great importance. Because instruction executions are partially, and not totally, ordered when parallelism is available, it is possible that resource requirements for executing a given program on a given set of data may vary from run to run. Running a program at one time may result in peak resource demands which exceed the machine's capacity (the computation cannot be completed), whereas running the same program another time may not.

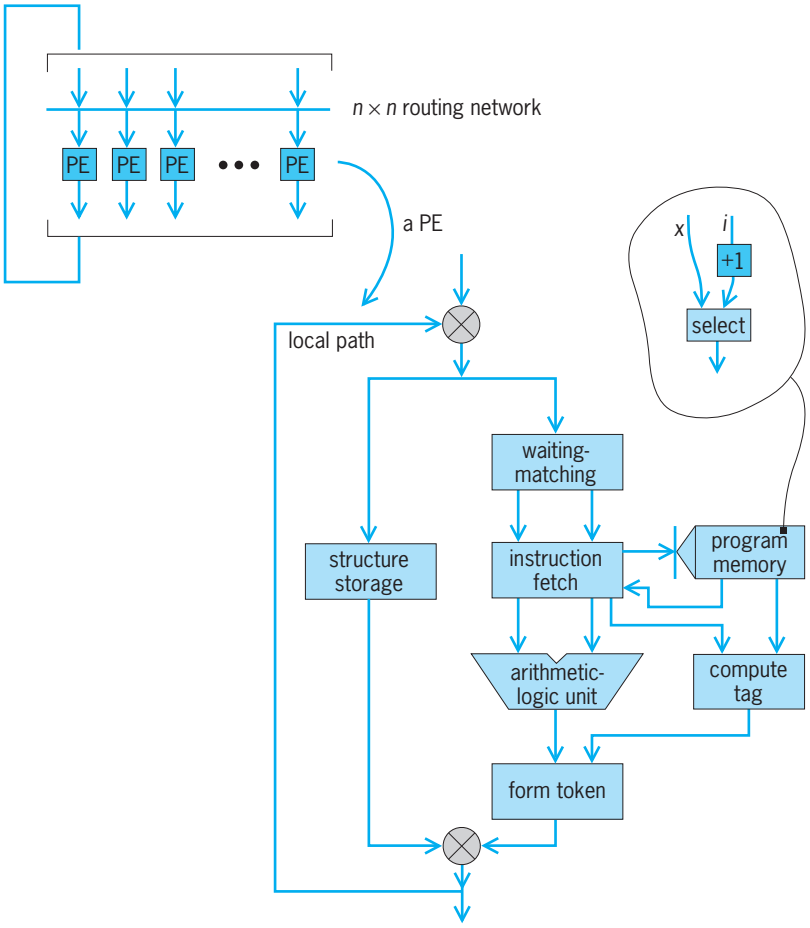

Fig. 2. Dynamic dataflow machine. PE = processing element. ***(After Arvind and R. A. Iannucci, in R. Dierstein, D. Müller-Wichards, and H.-M. Wacker, eds., Parallel Processing in Science and Engineering, Springer-Verlag, 1988)***

Known realizations of dataflow machines suffer from an inability to process efficiently sequential streams of instructions during periods of low-degree parallelism. This is attributable to the data-driven instruction execution rule in the presence of deep pipelines. Solutions to this problem involve combining conventional and dataflow-based instruction dispatching hardware. Such systems have come to be known as hybrid dataflow architectures. *See* COMPUTER PROGRAMMING; PROGRAMMING LANGUAGES.

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Date

The date palm (*Phoenix dactylifera*) is one of the oldest cultivated tree crops. It provides a staple food for many populations in the Middle East and North Africa, and also is highly valued for feed, fiber, and shelter. Beginning in the 1890s, the U.S. Department of Agriculture (USDA) plant collectors began importing date offshoots from Algeria, Tunisia, Egypt, and Iraq into the Salt River Valley in Arizona and then into the Coachella Valley in California, and the United States date industry was born. Date production and culture continue to challenge farmers around the world.

Botany. The date palm is in the family Palmaceae. *Phoenix* is distinguished from other palm species by the production of offshoots, columnar-circular trunks, pinnate leaves, and distinctively furrowed seeds. Date palms are monocotyledons and have a single bud or growing point. The date trunk also lacks the cambium growth layer typically found in dicotyledon fruit trees, which is important for true secondary growth.

The date palm species is dioecious: male palms produce staminate flowers and female palms produce pistillate flowers. Both male and female flower clusters are enclosed in protective sheaths. Bees and insects go to the male flowers but not to the female blooms. The date palm is dependent on humans for pollination. Pollen from the staminate flowers must be collected and manually applied to the pistallate inflorescence every season for commercial date production. Windblown pollination results in poor fruit set and in fruit that do not mature properly. One male tree can provide enough pollen for 30–40 female trees. Because the inner layer of the fruit wall is fleshy, the date fruit is classified as a berry, like a tomato.

Culture. The date palm grows under a wide range of temperatures, but fruit production requires 6000 to 7000 degree-days (Fahrenheit). Degree-days are the sum of the number of degrees of daily mean temperature above 50°F (10°C) during flowering and fruit ripening. In the Coachella Valley the medium temperature frequently exceeds the 110°F (43.3°C) and has been as high as 123°F (50.6°C). Temperatures below 20°F (−6.6°C) can cause injury to mature date palms; offshoots are injured at temperatures under 25°F (−3.9°C). Other requirements for date production are low rainfall and low humidity because prolonged periods of humidity affect developing fruit and reduce quality.

The date garden (a grower term) is planted with offshoots collected from established palms. Vegetative propagation is necessary because date seedlings do not duplicate the female parent, resulting in nonuniform fruit of inferior quality. Male palms should also be selected from offshoots for superior pollen source. Offshoots are connected to the parent tree by a short, woody base a few inches in diameter. Within the woody base is the terminal bud, it is protected by the lower leaf bases, and care must be taken to protect it from damage during offshoot removal. Offshoot size varies depending on variety, ranging 40–100 lb (18.1–45.4 kg) in weight and 8–14 in. (20.3–35.6 cm) in diameter. Commercial date plantings are spaced 30 by 30 ft (9.1 by 9.1 m) apart to provide sufficient sunlight and ventilation. Dates begin to fruit at around 5 years of age and are in full production at 10–15 years; they can remain productive for about 50 years. The root system is composed of long lateral roots that can extend as far as 20–30 ft (6.1–9.1 m) from the trunk, and secondary smaller roots. Root hairs, which are the main uptake of water and nutrients in most plants, are absent in date palms. Irrigation can be by borders or drip irrigation. Careful attention to irrigation is necessary to maintain palm growth and high fruit yields. Date palms can require up to 10 acre-feet (4 hectares) of water a year. Irrigation can be reduced during the harvest season in order to facilitate the harvest.

Fertilization. A total of 4–6 lb (1.8–2.7 kg) of actual nitrogen per palm is adequate in most soils. However, in light-textured soils where leaching is a concern, more nitrogen may be required. The nitrogen is used in split applications during the growing season. The source of the nitrogen can be from animal manures or winter or summer cover crops. High nitrogen applications may increase yields, but result in lowered quality.

Fruit ripening stages. Fruit thinning of bunches on the trees is necessary to increase fruit size, and it can be accomplished by reducing the number of fruits per bunch (bunch thinning) or by reducing the number of bunches per palm (bunch removal). The amount of quality fruit that a palm can safely carry depends on the age, size, vigor, and variety of the palm and on the number and size of green leaves that it has. Even after bunch thinning and removal, there may still be unequal ripening, within the bunch and with the other bunches on the tree.

Arabic terms are used in referring to the developmental stages of the fruit. "Kimri" refers to very young green-colored dates. The dates increase in size and weight and begin to rapidly accumulate reducing sugars. At this stage, dates begin to reduce in acidity and slightly gain moisture. In the "Khalal" stage the fruit just about reaches its maximum size and turns yellow, pink, red, or scarlet, or yellow spotted with red, depending on the cultivar. The "Rutab" or wet stage follow, and little or no sugar accumulates; the fruit continues to lose water. The fruit begins to soften and merges into the "Tamar" or cured stage; the fruit are fairly firm in consistency and have fully ripened. The sugar-to-water relationship in the fruit now prevents fermentation.

Grading. Each palm bunch must be hand-harvested. Field bins are transported to the packing houses for grading. The dates are cleaned and sorted into uniform lots and classes. The several lots comprise dates that are uniform in color, texture, size, and moisture content. Most packing houses have USDA inspectors during operating hours. Inspection covers adherence to grade standards, which include cultivar characteristics; uniformity of color; and freedom from insects, diseases, and various other defects; and size uniformity for the grade.

Varieties. In 1997 there were about 5640 acres (2282 ha) planted in California's Coachella Valley with an annual production value of $62 million. There are also 600 acres (243 ha) planted in Bard near Yuma, Arizona. The most popular of the many date varieties will be mentioned. The Deglect Noor cultivar is a semidry date and is the most popular variety in the United States; it is originally from Algeria. The Medjool is a soft date, originally from Morocco, that is prized for its extralarge size. The Zahidi is a semidry date from Iraq and has proved to be well adapted to California and Arizona. The Halawy is a soft date from Iraq, and its fruit is light amber to translucent golden brown when ripe. The Barhee, whose fruit is amber to deep golden brown on ripening, is also a soft date from Iraq. The Khadrawy is a soft date that is greenish amber on ripening and reddish brown when cured, and is also originally from Iraq.

Disorders. A common disorder is blacknose, a darkening and shriveling of the tip resulting in severe skin cracking on the fruit. During the growing stages the Banks grass mite causes serious drying and scarring of developing fruit. Fruit beetles can also attack ripe fruit, preferring soft and sour dates. These beetles lay eggs on the fruit, which the larvae feed on, making it unfit for consumption. If dates are not properly stored, fermenting yeasts and mold fungi decrease the quality. Physiological deterioration can also cause darkening and loss of flavor. Temperature, moisture content, and date variety affect this deterioration.

Pitted dates are now packaged for greater consumer acceptance and convenience. United States supermarkets stock pitted, whole, and chopped domestic dates. Because of the wide range of date textures available, packinghouses also produce a wide variety of products for use in the baking and institutional food-service industry. *See* ARECALES; FOOD ENGINEERING; FRUIT; FRUIT, TREE; INFLORESCENCE; REPRODUCTION (PLANT).

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Dating methods

Relative and quantitative techniques used to arrange events in time and to determine the numerical age of events in history, geology, paleontology, archeology, paleoanthropology, and astronomy. Relative techniques allow the order of events to be determined, whereas quantitative techniques allow numerical estimates of the ages of the events. Most numerical techniques are based on decay of naturally occurring radioactive nuclides, but a few are based on chemical changes through time, and others are based on variations in the Earth's orbit. Once calibrated, some relative techniques also allow numerical estimates of age. *See* ARCHEOLOGY; ASTRONOMY; GEOLOGY; RADIOISOTOPE.

Relative Dating

Relative dating methods rely on understanding the way in which physical processes in nature leave a record that can be ordered. Once the record of events is ordered, each event is known to be older or younger than each other event. In most cases the record is contained within a geological context, such as a stratigraphic sequence; in other cases the record may be contained within a single fossil or in the arrangement of astronomical bodies in space and time. The most important relative dating methods are stratigraphic dating and paleontologic dating. *See* SEQUENCE STRATIGRAPHY.

Stratigraphic dating. Stratigraphers rely on only a few basic principles to determine the order of events: (1) strata were horizontal (or nearly so) when initially deposited; (2) strata were laterally continuous over their original area of deposition; (3) younger strata lie on top of older strata; (4) a geological body must be younger than the particles of which it is composed; and (5) crosscutting features (such as faults and dikes) are younger than the rocks they cut.

Application of these principles allows determination of the order of events in local or regional geological sections and at archeological sites through determination of the physical relations between individual units. Exceptions and apparent exceptions to these principles are known but do not detract from their general applicability. Physical stratigraphic methods alone, however, do not allow the determination of relative ages of events recorded in sections that are widely separated geographically. *See* STRATIGRAPHY.

Paleontologic dating. Through physical stratigraphy, and close attention to the fossils contained within rocks, the principle of fossil succession was enunciated. This principle states that fossils succeed one another in a definite and determinable order, and that the period of geological time to which a rock body belongs can be recognized by the fossils included within it. Biological evolution is the fundamental process that makes this method work. Throughout geological history, different assemblages of organisms existed on the Earth at different times, and each organism has a definite time of appearance and (if extinct) disappearance. Paleontologic studies are the basis for the geologic time scale that is applicable to stratified rocks worldwide, for fossils allow correlation of rock units from one continent to another. The taxa present, the stage of evolution of related taxa, and the morphologic features within particular fossil taxa all change through time, and all may have temporal significance. Where quantitative techniques have been used to determine the temporal

range of a taxon, paleontological dating can be used to determine the age of rocks if direct quantitative methods cannot be applied. *See* FOSSIL; GEOLOGIC TIME SCALE; PALEONTOLOGY; SEDIMENTARY ROCKS.

Paleomagnetic dating. Throughout Earth history the magnetic field has changed its polarity, direction, and intensity. Most rocks contain sufficient iron oxide minerals to make them slightly or strongly magnetic, and they preserve a record of the orientation of the rock with respect to the Earth's magnetic field at that time. A principal use of these variations is chronological, and it is based on the fact that the polarity of the global magnetic field has reversed many times at highly irregular intervals. The timing of geomagnetic polarity reversals is reasonably well known from the Jurassic to the present. The present normal period does not differ from any that occurred in the past, so an initial estimate of the age of a rock sequence based on isotopic dates or paleontology is required before magnetic polarity zonation can yield additional chronological information. In a thick stratigraphic section where the rate of sedimentation was more or less constant, enough magnetozones may be encountered that their relative positions will allow a fit to the polarity-reversal scale with only rudimentary initial knowledge of the age. Note that certain rock types are susceptible to remagnetization, and spurious magnetozones may be encountered. Such magnetozones have no chronological significance because they arise from effects other than field reversal.

Scales of intensity and direction of the Earth's magnetic field have also been developed for many areas over the last few thousand years. These are principally of interest to archeologists (hence the term "archeomagnetism"), because pottery and bricks record the direction and intensity of the magnetic field at the time they were made. Thus, determining the direction and intensity recorded in such artifacts allows placement of the artifact in time once the scales of intensity and direction have been calibrated locally. *See* PALEOMAGNETISM.

Annual counting methods. In subtropical climates, trees produce a layer of spring wood and summer wood in response to the annual temperature cycle, thus making tree rings. In glacial lakes, sedimentation is controlled annually by the freeze-thaw cycle, so that a distinctive layer called a varve is produced. Counting back from the present, very accurate ages of tree rings can be determined, which have found use in calibration of the radiocarbon time scale. In geological sections that contain annual varves, precise estimates of duration of deposition are possible, even if the numerical age of the section is not known. *See* DENDROCHRONOLOGY; VARVE.

Tephrostratigraphy. Volcanic ashes from large silicic eruptions may be deposited in a very brief time over areas of subcontinental dimension. Each of these ashes is recognizable by the chemical composition of its glass phase and, once dated by quantitative techniques in one locality, allows a date to be assigned to a stratigraphic section wherever else it is identified. *See* VOLCANO.

Chemical methods. Several chemical processes occur slowly, producing changes over times of geological interest; among these are the hydration of obsidian, and the conversion of L- to D-amino acids (racemization or epimerization). Determination of age requires measurement of a rate constant for the process, knowledge of the temperature history of the material under study, and (particularly for amino acid racemization) knowledge of the chemical environment of the materials. *See* AMINO ACID DATING; OBSIDIAN; RACEMIZATION.

Natural volcanic glasses (obsidians) contain little water, but when a fresh surface is exposed through natural (breakage during transport; loading) or human (flaking) process, the surface absorbs water to form a hydrated layer, which thickens through time. With reasonable estimates of the temperature at which hydration occurs, and the composition of the obsidian, ages of obsidian artifacts and of geologically fractured obsidian can be obtained. The method has been applied principally in archeology, but also to obsidian fractured by glacial loading.

In biological systems, only left-handed (L-) amino acids are utilized, but right-handed (D-) amino acids are equally stable. In geological environments over time, there is a slow conversion of L- to D-amino acids (and back) until half of the molecules are right-handed. Thus, the ratio D/L changes from 0 to 1 (where only one asymmetric carbon is present in the amino acid). In theory, it should be possible to measure the age of ancient organic material by measuring the D/L ratio. However, the conversion process is very sensitive to temperature history and chemical environment, so the method has found use primarily as a stratigraphic tool. Few environments are constant in temperature over periods of several hundred thousand years, but the deep sea is one, and studies of deep-sea cores demonstrate that the method is useful in those environments.

While the bone-fluorine relationship is not strictly a dating method, it has some utility. Bones that have been buried and fossilized take up fluorine from ground water. Thus, older bones contain, in general, more fluorine than younger ones. By measuring fluorine, it may be determined whether an assemblage of bones is likely to be of the same age. Application of this method was instrumental in exposing the Piltdown hoax.

Calibrated isotopic scales. As with paleomagnetic polarity, other changes have occurred through Earth history that leave a record in the rocks. Among these are oxygen, carbon, and strontium isotopic ratios. Once the record of these changes has been calibrated by quantitative methods, measurement of the isotopic composition of materials in a geological section may date the material. The principal use of these methods has been in rocks of Tertiary or later age. *See* TERTIARY.

Rocks of different ages and rubidium content have different strontium isotope ($^{87}Sr/^{86}Sr$) ratios, and as they weather, the isotopic composition of strontium in the ocean changes. In the modern ocean, the $^{87}Sr/^{86}Sr$ ratio is uniform, and was presumably uniform in the past. Over the last 40 million years, there

has been a rise in the ratio of ^{87}Sr to ^{86}Sr in the oceans. As marine organisms build their shells of calcium carbonate, they incorporate small amounts of strontium, thus providing a record of the $^{87}Sr/^{86}Sr$ ratio in ocean water at the time they form. The $^{87}Sr/^{86}Sr$ ratio of shells or tests less than 40 million years old determines their age within about 1 million years. *See* STRONTIUM.

As water is removed from the oceans and locked up in glacial ice, the oceans become isotopically heavier because water ($H_2{}^{16}O$) evaporates more readily than heavy water ($H_2{}^{18}O$). Organisms that extract carbonate from the oceans record these changes, so that scientists were able to determine a curve of the isotopic composition of the oceans. As glaciers wax and wane, so does the isotopic composition of the ocean. Therefore, there are several times at which the isotopic composition of the ocean is identical. Thus, a good estimate of the age of specimens is needed before the oxygen isotope ($^{18}O/^{16}O$) ratio yields additional temporal information. However, knowing whether the $^{18}O/^{16}O$ ratio reflects glacial or interglacial times restricts the possible ages of a specimen to well-defined intervals.

Quantitative Methods

Unlike chemical methods, in which changes depend both on time and on environmental conditions, methods which are based on radioactive decay depend only on time. A parent nuclide may decay to one stable daughter in a single step by simple decay [for example, rubidium decays to strontium plus a beta particle ($^{87}Rb \rightarrow {}^{87}Sr + \beta$)]; to two daughters by branched decay through different processes [for example, potassium captures an electron to form argon, or loses a beta particle to form calcium ($^{40}K + e^- \rightarrow {}^{40}Ar$; $^{40}K \rightarrow {}^{40}Ca + \beta$)]; to one stable daughter through a series of steps (chain decay); or into two unequal-sized fragments by fission. In all cases, the number of parent atoms decreases as the number of daughter atoms increases, so that for each method there is an age-sensitive isotopic ratio of daughter to parent that increases with time. Many different isotopes have been exploited for measuring the age of geological and archeological materials. For example, the **table** shows parent isotopes, their half-lives, and the resulting daughter products. *See* HALF-LIFE.

Isotopic methods. All dating methods based on radioactive decay use simple equations (1)–(3) that

$$N = N_0\, e^{-\lambda t} \tag{1}$$

$$t = \frac{1}{\lambda}\, \ln \frac{N_0}{N} \tag{2}$$

$$t = \frac{1}{\lambda}\, \ln \left(\frac{D - D_0}{N} + 1 \right) \tag{3}$$

relate time (t) to the number of parent atoms originally present (N_0), the number of parent atoms that remain (N) after t, and the number of daughter atoms produced (D), accounting for any daughter atoms initially present (D_0); λ is the decay constant ($\log_e$ 2/half-life), and e is a fundamental-constant. To determine an age, N and D are measured, and D_0 is measured or estimated. In order for an age to reflect the time of the event in question, changes in the number of parent and daughter atoms must result only from radioactive decay. *See* RADIOACTIVITY.

Potassium-argon dating. Potassium is the dominant cation in micas, many alkali feldspars, and leucite. It occurs in high concentration in nepheline and many volcanic glasses, and in lower concentration in amphiboles, plagioclase, and basic volcanic rocks. The natural isotopic composition of potassium is constant; potassium-40 (^{40}K) decays either to calcium-40 (^{40}Ca) by beta emission or to argon-40 (^{40}Ar) by electron capture or positron emission. As a result, the $^{40}Ar/^{40}K$ ratio increases with time in potassium-bearing phases, and measurement of that ratio provides a measure of time. The method has been applied principally to igneous rocks and minerals, and has a useful range from about 100,000 years to the age of the Earth (4.5 billion years).

Argon is quantitatively retained by many solids at low temperatures, but diffuses rapidly through solid materials at high temperatures, so initial argon is not generally a problem. The event that is dated is the time of cooling to a temperature at which argon is effectively trapped, or the time at which low-temperature minerals crystallize.

In conventional potassium-argon dating, no initial ^{40}Ar is assumed to have been present in the sample. In most cases, this assumption is justified, but some samples have been shown to contain initial argon, and yield potassium-argon ages that are too old. Other samples have been shown to have lost argon, and yield ages that are too young.

In $^{40}Ar/^{39}Ar$ dating, a sample is first irradiated with neutrons to produce ^{39}Ar from ^{39}K. The ^{39}Ar content then becomes a measure of the potassium content of the sample. An advantage of this method is that the entire analysis is done on a single sample and not with separate aliquots for potassium and argon analysis. The age may be measured by extracting all of the gas at once, which produces an age comparable to that of a conventional potassium-argon determination. Alternatively, the gas may be extracted in a series of steps at higher and higher temperatures, and an age computed for each fraction of the gas; this method is referred to as step heating or incremental release. Ages are commonly plotted against temperature of release or the fraction of ^{39}Ar released, to

Principal parent and daughter isotopes used in radiometric dating

Radioactive parent isotope	Stable daughter isotope	Half-life, years
Carbon-14	Nitrogen-14	5730
Potassium-40	Argon-40	1.25×10^9
Rubidium-87	Strontium-87	4.88×10^{10}
Samarium-147	Neodymium-143	1.06×10^{11}
Lutetium-176	Hafnium-176	3.5×10^{10}
Rhenium-187	Osmium-187	4.3×10^{10}
Thorium-232	Lead-208	1.4×10^{10}
Uranium-235	Lead-207	7.04×10^8
Uranium-238	Lead-206	4.47×10^9

show a spectrum of ages. In this method, the gas released at low temperature comes from poorly retentive sites within the material, but gas released at higher temperatures comes from more retentive sites and therefore yields ages nearer the time of initial cooling. In an elegant variant of the method, a focused laser is used to heat a single crystal to extract the argon, from which an age is computed; this is called the single-crystal laser fusion $^{40}Ar/^{39}Ar$ method.

Rubidium-strontium dating. Rubidium-strontium dating has been widely applied to igneous and metamorphic rocks. The method is based on the beta decay of rubidium-87 (^{87}Rb) to strontium-87 (^{87}Sr), but because ^{87}Sr is present in significant amounts in most earth materials, it is essential that it be measured. This is accomplished by isotopic analysis of three or more cogenetic minerals or rocks, and plotting $^{87}Sr/^{86}Sr$ versus $^{87}Rb/^{86}Sr$. As ^{86}Sr is a nonradiogenic stable isotope of strontium, these ratios differ only because of differences in the original rubidium content of the material, and because of the formation of ^{87}Sr by radioactive decay. A plot of these ratios for an undisturbed system results in a straight line whose slope (m) is proportional to age [$t = m(e^{\lambda t} - 1)$], and whose y-intercept is the initial $^{87}Sr/^{86}Sr$ for the system. Ages measured by this method range from about 2 million to about 4 billion years.

Uranium-lead dating. Although they are chain decays, for ancient systems the decay of uranium and thorium isotopes to lead isotopes (^{238}U to ^{206}Pb, ^{235}U to ^{207}Pb, and ^{232}Th to ^{208}Pb) can be considered as occurring in a single step. Lead isotopic ratios are measured with respect to ^{204}Pb which is stable and is not a product of radioactive decay. In uranium- and thorium-bearing systems, the $^{206}Pb/^{204}Pb$, $^{207}Pb/^{204}Pb$, and $^{208}Pb/^{204}Pb$ ratios increase with time, and their measurement, together with measurement of the uranium and thorium content, allows computation of the time that has elapsed since the minerals crystallized. The methods are applied chiefly to zircon, a common accessory mineral in igneous and metamorphic rocks. If both the ^{238}U-^{206}Pb and ^{235}U-^{207}Pb systems have remained closed, both systems will yield the same age and are said to be concordant. The locus of all values of $^{206}Pb/^{238}U$ and $^{207}Pb/^{235}U$ that yield concordant ages is a line called concordia. Many ages are not concordant, and a series of determinations from the same rock may lie on a straight line (discordia) that leads from concordia to the origin, suggesting that the discordant ages are caused by recent loss of lead. In other cases, discordia intersects concordia twice, and may indicate both the time of original crystallization and a time of loss of lead in the past. For ages less than 200 million years, concordia is nearly a straight line, and Tera-Wasserburg diagrams are used instead, in which $^{207}Pb*/^{206}Pb*$ versus $^{238}U/^{206}Pb$ is plotted, where the asterisk implies radiogenic lead. These diagrams enhance the curvature of concordia for young ages. Uranium-lead methods are principally used on rocks older than 100 million years, but in recent years high-precision ages down to about 10 million years have been reported. *See* LEAD ISOTOPES (GEOCHEMISTRY); MASS SPECTROMETRY; ZIRCON.

Common lead dating. The time of isolation of lead from uranium-bearing systems into phases that contain no uranium can be determined by measurement of lead isotope ratios of sulfides, knowledge of the age of the Earth, and a reasonable estimate of the U/Pb ratio of the system from which the lead derived. This is called the common lead method, and has been applied to many ore bodies that are older than 100 million years in order to estimate the time of mineralization. *See* EARTH, AGE OF.

Uranium series disequilibrium methods. Many elements form as uranium decays by alpha and beta emission to lead. If a mineral containing uranium is closed to all elements in the decay series, each element will eventually be present in an amount related to its half-life. This condition is known as radioactive equilibrium, in which the activities (λN) of all elements are equal. Intermediate daughters may be separated from each other by geological or geochemical processes, leading to geochemical systems that are out of equilibrium. The quantitative departure of the activities from equality is a measure of the time that the series was disrupted. Several different isotopic pairs in the uranium decay series have been used for dating in the range from a few years to about 10^6 years. Most of these are in the ^{238}U decay series, but protactinium-231 (^{231}Pa) in the ^{235}U decay series has also proven useful for some systems. *See* PROTACTINIUM.

The principal use of uranium-disequilibrium dating systems has been in the study of deep-sea sediments; but some systems, notably $^{234}U/^{238}U$, have been applied to nonmarine carbonates with reasonable success. In water and minerals on continents the $^{234}U/^{238}U$ ratio is greater than unity, because ^{234}U is preferentially leached from minerals damaged by previous decays in the chain leading to it. Uranium is incorporated into calcium carbonate and other secondary minerals, and thereafter the excess ^{234}U decays more rapidly than it is formed, until the activity ratio returns to unity. The half-life for this process is 248,000 years. Another method for dating terrestrial carbonates is based on the $^{230}Th/^{234}U$ ratio. This is possible because calcium carbonate excludes thorium but includes significant quantites of uranium. As time passes, the amount of ^{230}Th increases because of the decay of both ^{234}U and ^{238}U in the sample. *See* DEEP-MARINE SEDIMENTS.

Other rock-dating methods. Many other methods based on decay of rare elements with long half-lives have been developed and applied since about 1980 to rocks older than about 10^7 years. Examples include the lutetium-hafnium (^{176}Lu-^{176}Hf) method applied to several common minerals, the rhenium-osmium (^{187}Re-^{187}Os) method applied principally on molybdenite and other sulfides, and the samarium-neodymium (^{147}Sm-^{143}Nd) method. Isochron techniques similar to those used in rubidium-strontium dating are normally applied.

Carbon-14 method. Radiocarbon dating is useful over the past 70,000 years and is based on the decay of

carbon-14 (^{14}C) to nitrogen-14 (^{14}N) by beta emission. ^{14}C is produced in the upper atmosphere by cosmic-ray bombardment of nitrogen, is oxidized to carbon dioxide, and enters biological systems through the photosynthetic reaction and metabolic reactions of herbivores; thus, living organisms are radioactive. Carbonates precipitated from water are also radioactive, because atmospheric radiocarbon exchanges with carbonate in solution in water bodies. When an organism dies or when carbonate minerals form, carbonates are removed from the active carbon cycle, and the ^{14}C within them begins to decay. To calculate an age, an initial activity (A_0) of ^{14}C is assumed, the activity (A) of the sample is measured, and ages are computed from Eq. (4), where

$$t = \frac{1}{\lambda} \ln\left(\frac{A_0}{A}\right) \tag{4}$$

t is time, and λ is the decay constant. Uncertainties in the age arise from analytical error, and from variations in the assumed initial activity. The latter are corrected by using curves calibrated against tree rings and by measuring ^{13}C content to correct for isotopic fractionation. Radiocarbon dates are most commonly measured on charcoal, wood, and carbonate minerals. *See* RADIOCARBON DATING.

Tritium-helium dating. This method is useful for dating ground waters, and depends on the decay of tritium (^3H) to helium (^3He); ^3H has a half-life of only 12.43 years, so the method is limited to water less than 40 years old. In this method, ^3He formed from decay of ^3H must be distinguished from ^3He derived from the atmosphere and other sources. A correction is made for atmospheric ^3He by measuring the concentration of dissolved neon; a correction for ^3He produced by subsurface nuclear reactions [primarily fission of lithium-6 (^6Li)] is made by measuring radiogenic ^4He in a sample. In this method, ^3H is normally measured by ingrowth of ^3He following complete degassing and a 6-month holding period. Tritium-helium ages have been used to evaluate ground-water recharge rates, timing of contaminant spills, ground-water flow rates, and as ground-water flow models. *See* TRITIUM.

Other methods based on radiation. Several methods of dating exploit the effects produced in crystals by particles emitted during radioactive decay. Thus, radioactivity is required, but trails of damage (fission track method) or trapped electrons (thermoluminescence, optically stimulated luminescence, and electron spin resonance methods) are measured rather than measuring the abundance of daughter nuclides, as is the case in the methods discussed above.

Fission track method. ^{238}U decays by alpha emission and also by spontaneous fission. The fission products disrupt the physical structure of the material in which the uranium is contained, and leave trails of damage that can be enlarged by chemical etching using strong reagents such as hydrofluoric acid or sodium hydroxide. The density of these fission tracks (number per unit area) is dependent on the age of the specimen and the uranium concentration; hence an age can be calculated by measurement of the track density and the uranium content. If materials containing fission tracks are heated, the tracks fade as the damaged parts are annealed. Thus, the event being dated is the time of cooling of the material below a temperature at which track fading is important.

To date a sample, a polished surface is prepared, the sample is etched, and the fission tracks are counted over a determined area. The uranium content is measured while irradiating the sample with a known dose of thermal neutrons that induce ^{235}U atoms to fission. As the $^{235}U/^{238}U$ ratio is known, the amount of ^{238}U may be calculated from the abundance of new tracks.

The most important mineral for fission track dating is zircon ($ZrSiO_4$), which contains relatively high concentrations of uranium and is very resistant to track fading. These qualities make it nearly ideal for production and retention of tracks. Sphene ($CaTiSiO_5$) has also been used, as have micas, but the latter have relatively low uranium contents and thus have low track densities in young rocks. Apatite [$Ca_5(PO_4)_3OH$] contains enough uranium to be useful, but its tracks anneal at low temperatures, and track fading is often a problem. Volcanic glass differs widely in its ability to retain tracks, but has been used for dating as well.

The method has been applied to artifacts as young as a few hundred years and to geological samples ranging from about 1 million to 1 billion years in age. *See* APATITE; FISSION TRACK DATING; MICA.

Thermoluminescence method. Some electrons freed from the valence bands of atoms by ionizing radiation from radioactive decay become trapped in crystal defects in many nonconducting materials. The number of such trapped electrons is a measure of the total radiation dose that a material has received; and if the annual dose rate is known, an age can be calculated by the equation t = dose/annual dose rate. When such materials are heated, electrons are freed from their traps, combine across luminescence centers, and emit light. The amount of light is a measure of the number of trapped electrons. The greatest difficulty with the method is determining the annual dose rate, but the method has found wide application in dating pottery, burned flints, and sedimentary deposits up to about 200,000 years old.

If light, rather than heat, is used to free electrons from their traps, the method is termed optically stimulated luminescence. *See* CRYSTAL DEFECTS; GEOCHRONOMETRY; THERMOLUMINESCENCE.

Electron spin resonance (ESR). As with thermoluminescence, this method measures the number of trapped electrons by measuring their absorption of microwave radiation. Paramagnetic centers with a static magnetic field are formed by traps occupied by a single electron. The energy of a microwave-frequency electromagnetic field applied to a sample is absorbed as the electrons in the traps oscillate between being oriented with and against the field, so measuring the amount of energy absorbed is a measure of the

number of electrons in traps. As with thermoluminescence dating, the greatest difficulty is in determining the annual dose rate. *See* ELECTRON PARAMAGNETIC RESONANCE (EPR) SPECTROSCOPY.

Cosmogenic nuclide dating. As cosmic rays interact with rock and soil at the Earth's surface, a variety of isotopes are produced by spallation, muon capture, neutron activation, and alpha-particle interaction. These isotopes include ^{3}He, ^{10}Be, ^{14}C, ^{21}Ne, ^{26}Al, and ^{36}Cl. The production rates are quite small, ranging from less than 10 to about 200 atoms per gram per year. Further, production rates vary with latitude, altitude, target composition, and time, and decrease rapidly with depth below the surface. The abundance of these cosmogenic nuclides in geological materials increases with time, and has been used to estimate surface exposure ages, erosion rates, and sedimentation rates. *See* COSMOGENIC NUCLIDE.

Astronomical Methods

These methods involve use of Milankovitch cycles to date various geological periods which have occurred on the Earth, and cosmochronology to determine the age of the universe.

Milankovitch cycles. Periodic changes in deposition in the deep sea are caused by small perturbations of the Earth's orbit around the Sun and the orientation of its rotational axis that result in small differences in energy flux to the Earth's surface, thus influencing climate. These changes occur at intervals of approximately 20,000, 41,000, and 100,000 years, at times reinforcing one another and at times opposing one another. The effects are expressed differently at different latitudes for cycles of different length. This technique has been used to refine the ages of the geomagnetic polarity time scale.

Cosmochronology. Astronomers have estimated the age of the universe, and of the Milky Way Galaxy, by various methods. It is well known that the universe is expanding equally from all points, and that the velocity of recession of galaxies observed from Earth increases with distance. The rate of increase of recession velocity with distance is called the Hubble constant; and knowing the recession rate and distance of galaxies at some distance, it is simple to find how long it took them to get there. Initial estimates for the age of the universe were approximately 20 billion years; but as the rate of expansion decreases with time, revised estimates are nearer 13 billion years. By contrast, the Earth and other bodies in the solar system are only about 4.5 billion years old. Comparison of present-day osmium isotope ratios with theoretically estimated initial ratios yields estimates of 8.6–15.7 billion years for the age of the Galaxy. *See* COSMOCHEMISTRY; COSMOLOGY; MILKY WAY GALAXY; UNIVERSE.

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Datolite

A mineral neosilicate, composition $CaBSiO_4(OH)$, crystallizing in the monoclinic system. It usually occurs in crystals showing many faces and having an equidimensional habit. It may also be fine granular or compact and massive. Hardness is $5–5\frac{1}{2}$ on Mohs scale; specific gravity is 2.8–3.0. The luster is vitreous, the crystals colorless or white with a greenish tinge. Datolite is a secondary mineral found in cracks and cavities in basaltic lavas or similar rocks associated with zeolites, apophyllite, prehnite, and calcite. It is found in the Harz Mountains, Germany; Bologna, Italy; and Arendal, Norway. In the United States fine crystals have come from Westfield, Massachusetts; Bergen Hill, New Jersey; and various places in Connecticut. In Michigan, in the Lake Superior copper district, datolite occurs in fine-grained porcelainlike masses which may be coppery red because of inclusions of native copper. *See* SILICATE MINERALS.

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Dawsoniidae

A subclass of the true mosses (Bryopsida) largely limited to the South Pacific. The subclass consists of a single genus, *Dawsonia*, of nine species. The gametophytes are remarkably similar to those of the Polytrichidae, but the origin and structure of the peristome are very different.

The plants are coarse and rigid in loose dark green or brownish tufts. They are simple or occasionally forked, grow erect from a procumbent base, and sometimes achieve a height of 26 in. (65 cm), much more than any other mosses. The stems may have a central strand of homogeneous fiberlike cells or both xylemlike and phloemlike cells. The leaves are long-linear or long-lanceolate from a relatively short, oblong-sheathing base. The limb, which may be as long as 1.6 in. (40 mm), is concave-tubulose and spreading or flat and spirally twisted when dry. It is coarsely toothed at the margins and at back. The costa occupies most of the limb, which is covered on its upper surface by 40–100 long green ridges, often with differentiated marginal cells. The cells of the sheath are long-linear, and those at the junction of sheath and limb are short and thick-walled. The inflorescences are dioecious and terminal. The perigonia are discoid with scalelike bracts. Male plants repeatedly form new growth through the perigonial bud. The setae are elongate and smooth or papillose. The

capsules are inclined, broadly ovoid-asymmetric, laterally compressed, and two-angled. The peristome teeth are long-filiform and somewhat twisted together in a bushy tuft. They consist of whole cells derived from eight concentric rows of amphithecial cells. They are united at the base into bundles of two to four U-shaped cells. The calyptrae are cucullate and matted with hairs. The haploid chromosome number is 7. *See* BRYOPHYTA; BRYOPSIDA; POLYTRICHIDAE.

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Day

A unit of time equal to the period of rotation of Earth. Different sorts of day are distinguished, according to how the period of rotation is reckoned with respect to one or another direction in space. A day is normally defined as 86,400 SI (Système International) seconds [86,400 s/d = (60 s/min) × (60 min/h) × (24 h/d)], where SI seconds are measured by atomic processes. *See* PHYSICAL MEASUREMENT.

Solar day. The apparent solar day is the interval between any two successive meridian transits of the Sun. It varies through the year, reaching about 24 h 30 s of ordinary clock time in December and about 23 h 59 min 39 s in September.

The mean solar day is the interval between any two successive meridian transits of an imagined point in the sky that moves along the celestial equator with a uniform motion equal to the average rate of motion of the Sun along the ecliptic. The difference between mean solar time and time measured by solar meridian transits is known as the equation of time. It leads to the analemma, the figure-eight seen on some globes. *See* EQUATION OF TIME.

Ordinary clocks are regulated to advance 24 h during a mean solar day. Because of irregularities in the Earth's rotation, leap seconds are occasionally added to keep the day in terms of seconds of atomic time (Coordinated Universal Time, or UTC) in step with the day based on rotation (UT1), though as of 2000 this system is under evaluation.

Sidereal day. The sidereal day is the interval between any two successive upper meridian transits of the vernal equinox. Similarly, as for the solar day, a distinction is made between the apparent sidereal day and the mean sidereal day which, however, differ at most by a small fraction of a second. The difference between sidereal time and solar time arises from the Earth's revolution around the Sun, which results in a year containing one more sidereal day than solar day. A mean sidereal day comprises 23 h 56 min 4.09053 s of a mean solar day.

The period of rotation of Earth with respect to a fixed direction in space is 0.0084 s longer than a sidereal day; the difference is caused by precession. No special name has been given to this kind of day, and although of theoretical interest, it is not used in practice.

Variations in duration. The mean solar day, the sidereal day, and the day mentioned in the preceding paragraph all vary together in consequence of variations in the speed of rotation of Earth, which are of three sorts: seasonal, irregular, and secular. The seasonal variations are probably caused, at least in part, by the action of winds and tides; the effect is to make the day about 0.001 s longer in March than in July, and is nearly repetitive from year to year. The irregular variations are probably the result of interactions between motions in the core of Earth and the outer layers; the effect is to cause more or less abrupt changes of several thousandths of a second in the length of the day, which persist for some years. The secular variation is the result of tidal friction, mainly in shallow seas, which causes the duration of the day to increase about 0.001 s in a century. *See* EARTH ROTATION AND ORBITAL MOTION; TIME.

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Dead reckoning

A form of navigation that determines position of a craft by advancing a previous position to a new one on the basis of assumed distance and direction moved. The name probably stems from the early practice of determining speed by throwing overboard a buoyant object, called a Dutchman's log, and noting the time needed for a known length of the vessel to pass the floating object, or attaching a line to the object (when the whole device became known as a chip log) and noting the amount of line paid out in a given time. In either case, the floating object was assumed to remain dead in the water, thus providing an indication of speed through the water. The reckoning of future positions of the vessel by means of this speed was known as dead reckoning.

Parameters. The parameters of dead reckoning are direction of motion and distance traveled. The intended direction of travel, the course, may differ from the direction steered because of the anticipated offsetting effect of wind (called leeway) or current, or both. When it is desired to distinguish between the two directions, mariners call the second the course steered, or heading, while aviators refer to it as the heading.

A compass is used to indicate direction. Distance is usually determined indirectly by measurement of speed and time, but it may be measured directly.

Measurement of direction. Direction is expressed as angular distance from a reference direction, usually north for courses and headings. This angular distance is customarily stated in integral degrees, from 000°

Fig. 1. Marine magnetic compass.

at north, clockwise to 360°. The reference direction used for advancing a position by dead reckoning is true north, the direction of the North Pole. The reference directions that may be used for courses steered and heading are magnetic north, the direction north along a magnetic meridian; compass north, the direction north as indicated by a magnetic compass; or gyro north, the direction north as indicated by a gyrocompass. For another reference direction, grid north, used in polar regions, *see* POLAR NAVIGATION.

Magnetic compass. For many centuries the primary direction instrument has been the magnetic compass. The first such compass probably consisted of a magnetic needle freely floated in water by means of a straw. This was replaced by the dry compass, and this in turn by the modern liquid compass in which the north-seeking magnetic element is placed in a compass bowl completely filled with a liquid which will not freeze at the temperatures in which the compass might reasonably be expected to operate.

For a typical modern marine magnetic compass (**Fig. 1**), the directional element, consisting essentially of several bundles of slender magnetized rods, is mounted by means of an almost frictionless bearing so that the compass will be responsive to weak magnetic fields.

Attached to the magnetic element is an annulus of lightweight material on which compass graduations are placed. This is the compass card, which remains aligned with the Earth's magnetic field as the vessel turns. Heading is indicated by the compass graduation aligned with a lubber's line in the longitudinal axis of the craft, generally on the forward side of the compass bowl aboard ship, and the after side in aircraft and land vehicles. Bearings can be measured by some compasses by means of a suitable attachment which is provided with sighting vanes.

Magnetic compasses are subject to compass error, the difference between true north and compass north. This error is the algebraic sum of two components. One of these, called variation by the navigator and magnetic declination by the magnetician, is the angle between magnetic and geographic meridians. It is related to the magnetic field of the Earth, and is subject to a small daily or diurnal change and to a slow progressive secular change. It is affected, also, by magnetic storms.

The second component is deviation, the angle between the magnetic meridian and the axis of the compass card. It is the result of local magnetic influences, particularly those within the craft. Permanent magnetism in metal of the craft, and transient magnetism induced there by the Earth's magnetic field are important contributors. Direct currents in electrical wiring also influence the compass. *See* MAGNETIC COMPASS.

The various disturbing influences can be largely neutralized by establishing additional magnetic fields of equal strength but opposite polarity. This process is called compass adjustment, or sometimes compass compensation, especially with respect to aircraft compasses. In the U.S. Navy the latter expression is used to indicate the process of neutralizing the effects that degaussing currents exert on a magnetic compass. These currents may be used to change a ship's magnetic characteristics to protect against magnetic mines.

In some compass installations, particularly in aircraft, the magnetic compass is installed in a position relatively free from magnetic disturbances, and its indications are transmitted electrically to locations throughout the craft. Such an instrument is called a remote-indicating compass.

Compass error is determined by comparing the compass with one having a known error, with the azimuth of a celestial body, or with the bearing of a landmark.

When deviation is reduced to a minimum, the residual deviation on a number of headings is recorded on a deviation table or deviation card (**Fig. 2**) kept near the compass.

The directive force of the compass is the horizontal component of the Earth's magnetic field at the compass. This becomes progressively weaker as

#2 MASTER COMPASS
SWUNG: 7-24-58 BY: KB

TO FLY	STEER	TO FLY	STEER
N	359	180	181
15	15	195	196
30	30	210	211
45	44	225	226
60	60	240	241
75	75	255	256
90	90	270	271
105	105	285	286
120	121	300	301
135	135	315	316
150	151	330	330
165	166	345	345

Fig. 2. Typical aeronautical deviation card.

the magnetic poles are approached. Within a few hundred miles of these poles the magnetic compass is unreliable. *See* GEOMAGNETISM.

Gyrocompass. In addition to several magnetic compasses, nearly all naval vessels and ocean liners are equipped with one or more north-seeking gyrocompasses. This instrument has the disadvantage of requiring a source of electrical power, but it is not subject to variation or deviation. It may have a small gyro error, but in most modern installations this is not large enough to be significant, except in high latitudes. The compass tends to align its rotational axis with that of the Earth but, since it is constrained to remain in the horizontal, its directive force decreases as the horizontal becomes more nearly perpendicular to the Earth's axis. Beyond some latitude, generally about 70°, the gyro error becomes large and erratic and the instrument requires frequent checking. Near the geographical poles the gyrocompass becomes unreliable.

Great progress has been made in the development of better gyroscopes. As a result, gyrocompasses have become smaller, lighter, more accurate, and practical for smaller vessels. Gyrocompasses have replaced magnetic compasses as the primary source of directional information on many modern vessels. *See* GYROCOMPASS.

Directional gyro. A directional gyro, used in some aircraft and submarine compass systems, tends to maintain a fixed position in a great-circle plane. Better instruments are precessed to allow for rotation of the Earth, so that a craft following their indications tends to follow a great circle on the Earth. Since the instrument does not seek anything, it is subject to a slow, cumulative error called wander. This is allowed for or corrected at intervals by comparison with a directional reference, generally a magnetic compass or the azimuth of a celestial body. *See* AIRCRAFT COMPASS SYSTEM.

Several types of devices provide directional guidance in polar regions.

Distance or speed measure. Distance is generally stated by navigators in units of nautical miles, usually to integral miles by air navigators and to tenths of a mile by marine navigators. For this and other linear units of navigation *See* NAVIGATION.

Aboard ship, distance or speed is measured by means of a log or by an engine revolution counter. The primitive Dutchman's log and chip log were mentioned earlier. The latter is of particular interest because a series of knots was tied in the log line, and the number of knots which passed through the seaman's hands in a given time interval represented the speed of the vessel in nautical miles per hour. This is the origin of the name knot applied to the unit of speed still commonly used by marines.

About the middle of the seventeenth century, mechanical logs appeared. These had rotators towed through the water, first with the indicator also in the water, but later at the taffrail, giving the name to the taffrail log widely used for many years.

The pitometer log uses a pitot-static tube. The Forbes log uses a small rotor in a tube projecting below the bottom of the vessel. An electromagnetic log has a sensing element which produces a voltage directly proportional to speed through the water. Accurate so-called measured miles are marked out at various places on the beach to provide means for calibrating or checking logs.

In aircraft, speed through the air is measured by means of an airspeed indicator or Mach meter. The latter provides an indication of speed in units of the speed of sound, which varies with density of the atmosphere. For measurement of air speed a pitot-static tube is generally used with a suitable registering device. Distance is determined indirectly from speed and time. Distance, and indirectly speed, may also be determined by means of two fixes, the length of the line between them being the distance.

Practice. To determine position by dead reckoning, air and land navigators, and some marine navigators, use the best estimate of direction and distance traveled over the surface. Many marine navigators, however, prefer to use course steered and estimated speed through the water, without allowance for leeway, for their dead reckoning; they consider positions determined by allowance of estimated effects of wind and current as estimated positions.

The uncertainty of a dead-reckoning position, however determined, increases with time and, if there is an error in direction measurement, it also increases with distance traveled. From time to time an independent determination of position is made by means of external references. When a reliable position, called a fix, is so obtained, a new dead reckoning is started from this point.

The total combined effect of all errors since the last fix can be determined by comparing the new fix with the position that the craft would have occupied if there had been no errors. It is for this reason, in part, that many marine navigators prefer to use course steered and speed at which a vessel is propelled by its primary source of power for their dead reckoning, thus perpetuating the original concept of reckoning relative to an object dead in the water (motionless with respect to the water). For convenience, they consider the entire offset effect as the result of current. The true direction from the dead-reckoning position to a fix at the same time is called the set of the current. The distance between these two positions in nautical miles, divided by the time in hours since the last fix, is called the drift (speed) of the current in knots. Air navigators consider the entire offset the result of wind, determined by comparison of a no-wind position, also called an air position, with a fix at the same time. The no-wind position is the point at which the aircraft would have arrived if there had been no offsets.

Determination of position. Several different methods have been used for determination of position by dead reckoning.

Computation. Before reliable charts became generally available, dead reckoning was commonly performed by computation using various forms collectively known as sailings. This method is seldom used

by modern navigators except in small craft where chart work is difficult, or when computation of direction and distance between widely separated points is desired. A computer or hand-held calculator may be used to make a computation. *See* CALCULATORS; DIGITAL COMPUTER.

Plotting. As reliable charts and plotting sheets came into general use, plotting became the common method of performing dead reckoning, a method still widely used. Marine navigators usually measure the plotted direction by means of parallel rules (a device designed to move over the chart or plotting sheet parallel to itself) and a compass rose, or a drafting machine (which combines parallel motion with direction indication). Dividers are usually used with the latitude graduations on the chart or plotting sheet for measuring distance. Air navigators usually use some form of plotter, one type of which is shown in **Fig. 3.**

Dead-reckoning computers. In many large naval vessels and commercial ships the dead reckoning is performed automatically by a device that receives inputs of direction from a gyrocompass and speed from a log and continuously computes dead-reckoning position, which is displayed on dials or traced on a chart or plotting sheet.

Doppler navigation. The Doppler effect, a frequency shift that is proportional to the speed of relative motion between transmitter and receiver or reflector of radiant energy, either acoustic or electromagnetic, is used in a system to accomplish dead reckoning, automatically. In the ship version, called a Doppler sonar navigator, ultrasonic energy is transmitted obliquely downward (typically 30° from the vertical) and the frequency of the return echo is noted. By using four beams separated 90° laterally, the system provides an indication of vessel speed in both the fore-and-aft and athwartship directions, so that total speed and direction of motion can be determined if the device is properly oriented. When reflections are from the sea bottom (bottom return mode), true ground speed (speed relative to the solid Earth) is measured. When reflections are from suspended particulate matter in the water (volume reverberation mode), the speed is relative to the water, similar to speed relative to an object dead in the water. In either mode the speed is integrated to determine distance from a starting point. Doppler sonar navigation has proved particularly useful in survey and geophysical exploration vessels. *See* SONAR.

Similar systems called Doppler navigators, which use electromagnetic energy, have been used in aircraft, but these systems have largely been replaced by inertial navigators. *See* DOPPLER EFFECT; DOPPLER RADAR.

Inertial navigation. Most aircraft that are used for long overwater flights, and some others, as well as some ships, notably submarines, are equipped with one or more inertial navigators. This device, when properly aligned, provides a continuous indication of speed, position, and heading by means of appropriate inertial sensors. Gyroscopes are used to sense angular motions of the craft and maintain accelerometers in the correct orientation to sense linear accelerations, or changes in speed. Single integration of the accelerations provides a measure of speed, and double integration produces a measure of distance. The term inertial stems from the fact that measurements are made with respect to inertial space. Provision must be made to eliminate effects related to the rotation of the Earth because data relative to the Earth are desired.

An inertial navigator is self-contained, because it is independent of external reference points, and passive, because no energy is transmitted to obtain information from an external source. It is free from effects of wind and current, but like all dead-reckoning systems its output degrades with elapsed time and distance traveled, and must be reset periodically. Its use is particularly attractive in aircraft because of their high speed and hence relatively short time in transit. *See* INERTIAL GUIDANCE SYSTEM.

Continuous position indication. Knowledge of the real-time present position of a craft is generally considered essential to safe navigation. Dead reckoning, always available in some form, provides this capability, giving information useful in the evaluation of all other navigational data received. Satellite navigation systems such as the Global Positioning System can provide an essentially continuous position fix when enough satellites are in view. An integrated navigation system that combines satellite observations with measurements from inertial instruments, Doppler sonar, or other navigation sensors can use this redundancy to maintain accurate positioning during brief intervals of satellite outages. *See* CELESTIAL NAVIGATION; ELECTRONIC NAVIGATION SYSTEMS; PILOTING; SATELLITE NAVIGATION SYSTEMS.

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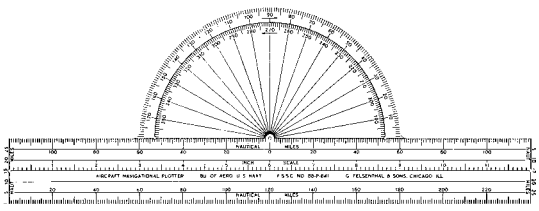

Fig. 3. A plotter in use among air navigators.

Death

Cessation of life functions. This can involve the whole organism (somatic death), individual organs (organ death), individual cells (cellular death), and individual parts of cells (organelle death). Although the next smaller level or organization, the macromolecules that make up the cell organelles, may also cease to function, their disintegration is ordinarily not spoken of as death.

Pathogenesis

Figure 1 represents a useful conceptualization of the dynamics of injury, death, and necrosis. A normal animal, cell, or organelle can be assumed to exist in a normal state of homeostasis or a normal "steady state." Following a disturbance of the normal steady state, referred to as an "injury," the living system can be assumed to undergo a change in the steady-state level, that is, to exhibit an altered steady state. Such altered steady states can be regarded as having increased, decreased, or unaltered levels of ability to maintain homeostasis. An altered steady state with increased homeostatic ability might include the hypertrophy of a muscle cell resulting from continued exercise; an altered steady state with decreased homeostatic ability is the atrophy or shrinking of cells that occurs following disuse or lack of stimulation by hormones or suitable nutrients. In the case of injuries that do not result in the death of the living system, the system may persist in one of the new or altered steady states for long periods up to many years; that is, although the system differs significantly from the normal steady state, it is able to continue its existence and to maintain homeostasis. On the other hand, certain severe injurious stimuli result in an altered state that ultimately passes a point beyond which recovery of homeostatic ability is impossible even if the injurious stimulus or situation is removed. This may be considered as a point of no return or the point of death of the living system. Following this point, the system undergoes exponential decay, leading to complete equilibrium with its environment. This period of decay, which is characterized by reactions resulting in the breakdown of structural components of the system, is referred to as necrosis. *See* HOMEOSTASIS.

Somatic death. Somatic death has different meanings for the biologist, the theologian, the physician, and the lawyer. In the biological sense, somatic death refers to the cessation of characteristic life functions. In higher organisms, death can be assumed to result from an interplay between the genetic background of the individual and the effects of the environment. Death is, in a sense, the price of differentiation; in another sense, protoplasm is immortal and all life springs from preexisting life. The undifferentiated amebas need never die, only divide. It would appear that death is inevitable for higher organisms such as mammals and, in a sense, life is a series of little deaths.

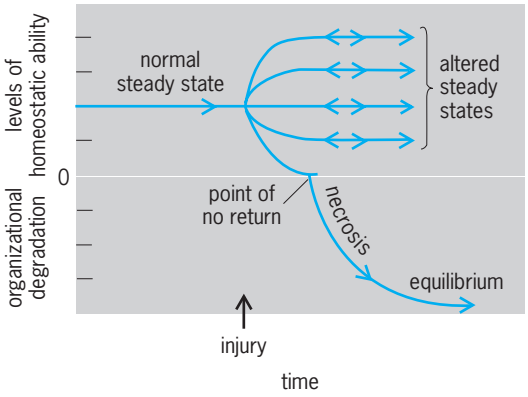

Fig. 1. Changes that can occur in living systems following injury. In the case of lethal or fatal injuries, the system can eventually be assumed to pass a point, shown as zero level of homeostatic ability, beyond which it is unable to restore a normal state of homeostasis even if the injurious situation is removed. (*After E. Bajusz and G. Jasmin, eds., Methods and Achievements in Experimental Pathology, vol. 4, Karger, Basel, Switzerland, 1969*)

Forensic medicine. Developments in medicine and associated technology involving supportive therapy, resuscitation, hypothermia, extracorporeal circulation, and organ transplantation reopened serious questions regarding the definition of when an individual is "dead." The vagueness and inconsistency in previous legal and medical definitions have led to serious dilemmas. These are of particular importance in the selection of donors for transplantation of vital organs such as the heart. It has been proposed that death be defined as irreversible cessation of all of the following: (1) total cerebral function, including determination by electroencephalography; (2) spontaneous function of the respiratory system; and (3) spontaneous function of the circulatory system. *See* FORENSIC MEDICINE.

Technology and prolonged survival. Medical techniques have shown that people can be sustained for prolonged periods without spontaneous respiratory or cardiac movements as well as without electrical evidence of central nervous system activity. These techniques have also achieved the resuscitation of individuals who would formerly have been considered dead and have clearly shown the survival of transplanted organs from cadavers.

Because of the present ability to maintain cardiac and respiratory function, even in the presence of complete absence of cerebral activity, special emphasis has been placed on laboratory techniques such as electroencephalography, coupled with clinical evaluation, to determine whether irreversible cessation of cerebral function has, in fact, occurred. If intravital brain death can be fully established clinically as well as by the electroencephalogram, further attempts to maintain the patient, even if successful, result only in incubation of a "dead" brain at the body temperature of the patient. Just as organs may survive after the death of the organism, the organism, properly supported by mechanical or other means, can survive the death of organs, including the brain.

Postmortem changes in body. The following are changes that occur in the body after somatic death. The rate of progression and time of occurrence of these changes are of some use in determining the time and circumstances of death and, accordingly, are important in forensic medicine.

1. *Cooling (algor mortis).* This refers to the cooling of the body that occurs postmortem. The rate of algor mortis is affected by environmental temperature, conductivity of the environment, amount and type of clothing, and amount of body fat.

2. *Rigidity (rigor mortis).* This refers to the stiffening of the musculature. Although its time of onset is somewhat dependent on the degree of muscular

activity at the time of death, it usually begins within 5–10 h. The muscles of the jaw are first affected. Rigor mortis begins to disappear after 3 or 4 days.

3. *Staining (livor mortis)*. This refers to the reddish-blue discoloration that occurs in the dependent portions of the body and results from the gradual gravitational flow of unclotted blood.

4. *Clotting of blood*. Blood begins to clot shortly after death and sometimes prior to cessation of cardiac function. Such clots are not attached to the vessel wall and can be distinguished by this, as well as by their appearance, from antemortem thrombi.

5. *Autolysis*. This refers to the necrosis that occurs in the cells of the body following somatic death. The changes are indistinguishable from those occurring following ischemic (deprivation of blood supply) injury in the body, although an inflammatory reaction, a response to necrosis in tissues of the body, is absent.

6. *Putrefaction*. This refers to the bacterial and enzymatic decomposition that begins shortly after death, caused by microorganisms, including bacteria and fungi that enter the body, usually from the gastrointestinal tract. It results in production of gases and a greenish discoloration of the tissues.

Organ death. Following somatic death, all organs do not die at the same rate. Thus, certain vital neurons in the brain survive for approximately 5 min, the muscle cells in the heart for approximately 15 min, the kidney cells for about 30 min, and the liver cells for 45 min. If a blood supply and a suitable environment are restored to these organs prior to the point of no return, they will recover normal function. Length of survival can be prolonged by reducing the temperature of the organs to near 32°F (0°C), which appreciably retards metabolic function and extends survival time. This survival of organs after the death of the host forms the basis for organ transplantation from cadavers.

Cellular death. Death and subsequent necrosis of cells form an important part of many reactions of cells to injury in disease states. Thus, in the heart after a coronary occlusion, the cells of the myocardium die and undergo disintegration. In this case, the physiological consequences, cardiac failure and death of the organism, are the direct consequences of the death and disintegration of individual muscle fibers. The death and disintegration of cells within a living animal incite a vascular reaction called an inflammatory reaction on the part of the host. It is characterized by a series of vascular phenomena which include the migration of leukocytes (white blood cells) from the blood capillaries into the area of tissue damage with subsequent digestion and phagocytosis of the cellular debris resulting from cellular necrosis.

Cellular changes. The essential features of necrosis are similar following a variety of types of lethal injury to cells. Thus, the changes in the necrotic heart muscle cells mentioned above are essentially similar to those in the neurons of the brain after a stroke, to those in the cells of the lung in a tuberculous infection, and to those of the liver cells during viral hepatitis. The cell and its organelles show a characteristic series of changes involving the plasma membrane, the mitochondria, the endoplasmic reticulum, the lysosomes, and the nucleus (**Figs. 2** and **3**).

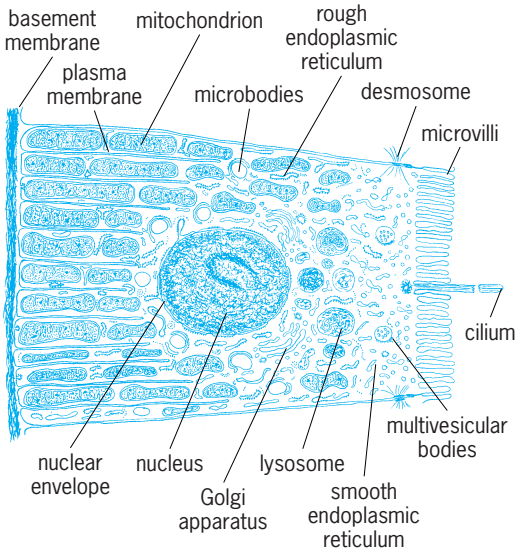

Fig. 2. Diagrammatic representation of a normal kidney tubule cell as seen with the electron microscope. (*After F. L. Ginn, J. Shelburne, and B. F. Trump, Disorders of cell volume regulation, I: Effects of inhibition of plasma membrane adenosine triphosphatase with ouabain, Amer. J. Pathol., 53:1041–1047, 1968*)

In a normal kidney tubule cell, at the apex of the cell adjacent to the tubular lumen, the plasma membrane forms numerous microvilli; at the base of the cell, the plasma membrane is highly invaginated and forms complex compartments that contain the elongate mitochondria. The mitochondria are limited by two membranes; the inner membrane forms a series

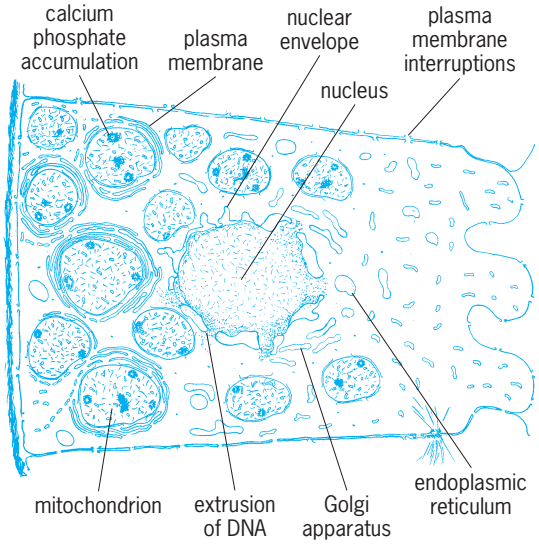

Fig. 3. Diagram showing a necrotic kidney tubule cell. (*After F. L. Ginn, J. Shelburne, and B. F. Trump, Disorders of cell volume regulation, I: Effects of inhibition of plasma membrane adenosine triphosphatase with ouabain, Amer. J. Pathol., 53:1041–1047, 1968*)

of invaginations called cristae, which are responsible for production of adenosine triphosphate (ATP). The nucleus, in the center of the cell, is surrounded by the double-walled nuclear envelope, the lumen of which is continuous with that of the endoplasmic reticulum. The endoplasmic reticulum has two portions, smooth- and rough-surfaced. The rough-surfaced endoplasmic reticulum is covered with particles called ribosomes, on which synthesis of protein for secretion from the cell occurs. The Golgi apparatus is in the supranuclear region. Small organelles called microbodies appear to be responsible for a primitive form of cellular respiration.

A characteristic feature of necrotic cells includes massive swelling of the cytoplasm, often referred to in the past as cloudy swelling. This swelling results from movements of ions, such as sodium and water, into the cell as a result of the loss of active transport mechanisms at the cell membrane. The swelling involves specific organelles such as the endoplasmic reticulum and the mitochondria. The endoplasmic reticulum becomes greatly dilated, and the ribosomes along the surface of the reticulum are disorganized, with loss of ability to synthesize protein. The plasma membrane is whorled near the base of the cell and becomes interrupted elsewhere. The mitochondria exhibit massive swelling of their inner compartment, and in this stage have lost the ability for production of ATP. Mitochondria also contain accumulations of calcium phosphate. The nucleus undergoes lysis, often called karyolysis, and the nuclear deoxyribonucleic acid (DNA) extrudes into the cytoplasm through the nuclear pores. The lysosomes in this stage rupture, with release of digestive enzymes into the cytoplasm. Such enzymes are evidently responsible for the disintegration of cell components that occurs during the necrotic phase.

Subcellular changes. The elucidation of the myriad of subcellular changes that occur in injured, dying, and dead cells is possible through the use of the electron microscope and associated techniques. With this instrument, which is capable of resolutions much greater than those obtainable by light microscopy, a close correlation between the structure of organelles, or parts of organelles, and the physiological and biochemical functions of these same structures is possible. Use of the electron microscope permits the demonstration of cellular damage at very early intervals following application of an injury. Such studies were impossible by light microscopy alone, and in the past it was necessary to wait for many hours for the characteristic changes to develop following death of the cell. The normal, the injured, and the dead cell are difficult to delineate clearly and exist along a continuum of cell structure and function.

Biochemical parameters. It is believed that the limiting factors in the survival of most cells following a lethal injury are the integrity of the plasma membrane and interference with ATP synthesis by mitochondria or other organelles. Although the lysosomes, which are membrane-bound vacuoles containing hydrolytic enzymes, are sometimes thought of as "suicide bags" which may kill the cell by premature release of enzymes, most experiments indicate that release of hydrolytic enzymes from these bodies does not occur until after cellular death, at which time such release catalyzes hydrolytic reactions resulting in disintegration of cell components. *See* LYSOSOME.

Surviving elements. After death of the cell as a unit, some parts survive longer than others. Just as the heart may continue to beat for a short time after respiration has ceased, protein synthesis within the endoplasmic reticulum may occur after cessation of respiration in mitochondria. Because of the great complexity of changes that can occur within the cell, and because of the dissociation between their time of occurrence, an exact definition of cell death is extremely difficult.

Organelle death. Following the death of the cell, organelles continue to function for variable periods of time, depending upon the manner in which death occurred. Thus, in cells dying from ischemia, the mitochondria continue to respire for an hour or more, although the synthesis of ATP is impaired. Also, portions of the endoplasmic reticulum exhibiting polyribosomes (rough-surfaced endoplasmic reticulum) may continue to synthesize protein. The lysosomes maintain their enzyme activity for as long as 2 days following death of the cell at 98.6°F (37°C). Their enzymes are thought to be important in the removal of the products of dead tissue by hydrolytic digestion.

The survival of organelles after the death and disintegration of the cell is well illustrated by experiments in which cells are destroyed by mechanical homogenization. In such experiments, cell organelles, including fragments of the endoplasmic reticulum or mitochondria, can then be maintained for relatively long periods of time in the test tube. It has been shown that fragments of the endoplasmic reticulum can continue to synthesize protein following disintegration of the cell and that mitochondria can respire, synthesize ATP, and even continue to replicate after the cell has been utterly destroyed. Such experiments illustrate the relative independence of organelles such as mitochondria and even suggest the possibility of organelle transplantation.

Autophagocytosis. During the life of most, if not all, cells, portions of the cytoplasm are enclosed from time to time in vacuoles that come to contain hydrolytic enzymes. These vacuoles are known as autophagic vacuoles or autolysosomes. Death and disintegration of the enclosed organelles occur in these vacuoles. Such autophagic vacuoles may be an important site of normal organelle turnover and may provide the cell with energy sources in times of nutritional deficiency or starvation.

Necrobiosis. This is a term used to indicate physiological death of cells and their replacement within the living organism. Examples of death and turnover of cells in normal living animals are seen in aging, in shedding of the superficial cells of the epidermis of the skin or the epithelium of the gastrointestinal

tract, and in the involution of the breast and uterus following pregnancy. Benjamin F. Trump

Issues

The advancements of modern scientific medicine have resulted in significant changes in the care and treatment of individuals with life-threatening illness, and challenge concepts and definitions of death. Considerable efforts are being made toward establishing clarity by developing a more precise definition of death and the corresponding criteria for determining death. These changes are accompanied by a cluster of persistent ethical issues such as the right to refuse life-saving treatment, the circumstances under which it is morally right not to prolong life, and whether extreme situations justify a deliberate act of euthanasia or physician-assisted suicide. Legislation, court decisions, and public referendums are changing the shape of public policy surrounding death and dying in the United States, even though a clear moral consensus may be difficult to reach.

As the determination of death becomes increasingly complex in a technological society, the definition of death becomes correspondingly more complicated. The meaning of death is a philosophical concept which lies outside the scientific point of view. Traditionally, death has been understood by the general public, and by the physician as well, as the cessation of life or ceasing to exist. Although dying is a process that may continue over a considerable period of time, for all practical purposes death occurs at that point where there is an irreversible loss of one's essential human characteristics. Such loss indicates the death of the organism as a whole, the death of a person.

Historical perspective. Up until the seventeenth century in western civilization, death was considered a discrete moment in which the soul left the body. This point was determined by the absence of visible manifestations of life such as were apparent to the ordinary person, absence of breathing, and cessation of circulation.

In the eighteenth and nineteenth centuries a concern arose over premature burial as people began to doubt the accuracy of judgment of those who made the assessment that death had occurred. During this period life signals were installed on caskets with the belief that a person mistakenly believed to be dead should have an opportunity to sound an alarm until burial had been accomplished. Laws arose requiring physicians to examine the body prior to burial, and interment was sometimes delayed until the signs of putrefaction proved death undeniably. In some cases, boiling liquids or surgical incisions were administered as final proof.

In earlier times, when heart or lung function ceased, the brain also died within a few moments. Conversely, when the brain died, heart and lung function ceased as well. Today, however, the capabilities of modern medical technology to support or replace failing organ functions has completely changed the inevitability of this cause-effect relationship.

In the past, there was no evidence of concern for a condition in which persons appearing to be alive were in actuality dead. On the contrary, the concern was uniformly for false positives. Yet it is just this concern, that of the false negatives, which is being examined by physicians, philosophers, theologians, and lawyers. The general public has also demonstrated growing concern for this issue.

There are four major factors which seem to demand the more precise definition of death: (1) modern medicine's technological ability to sustain life in the absence of spontaneous heartbeat or respiration; (2) the advent of successful organ transplantation capabilities, which creates a demand for viable organs from recently deceased donors; (3) the enormous expenditure of resources which would be wasted if persons who in actuality are dead are being treated medically as though they are alive; (4) respect for persons requiring that the distinction be made when death has occurred, and that the dead person is treated appropriately in the light of this change in status.

New concepts. Throughout history, the concept of death has been tied to one or possibly two factors: the loss of the soul from the body, or the loss of vital-fluid circulation in the body. Efforts to establish a more precise definition of death have moved beyond the spontaneous function of the respiratory and circulatory system. The body's capacity for integrating its function is identified as the essentially significant indication of life. In the suggested new concept of death, it is not the death of one or more organs which is definitive, but the death of the whole person. When the point is reached in which these essential integrative characteristics are irreversibly lost, death occurs.

A still further refinement is offered by those working toward a more precise definition. The human capacity for integration can be divided into two categories. The first is the integration of the internal bodily mechanisms which are governed unconsciously, such as respiration and the regulation of blood pressure. The second is the integration of the person with the social environment through consciousness. This capacity for consciousness permits self-reflection and self-awareness as well as social interaction with others. Thus, this further refinement would claim that when the irreversible loss of the capacity for consciousness occurs, that which is essential to life is lost and the person is then dead. An even more radical concept is that death is the irreversible loss of the capacity for social interaction. *See* CONSCIOUSNESS.

Locus of death. Depending upon the philosophical concept of death that has been accepted, it becomes more specifically apparent as to where to look to determine the loss of the essential characteristics of human life. Thus, in terms of a vitalist concept, death is the loss of the circulation of vital fluids, and it would be appropriate to examine and evaluate the function of the heart and lungs. If the concept of integrative function is accepted as being the essential characteristic, it would be appropriate to evaluate the total brain. If, on the other hand, death is the loss of the embodied capacity for social interaction,

then all evidence today points to the neocortex as the locus of consciousness and social interaction.

New criteria. Whereas the concept of death is a philosophical issue, the determination of the irreversible loss in the relevant locus of the body is primarily the task of the biomedical sciences.

An interdisciplinary committee at the Harvard Medical School accepted the task of examining possible new criteria for determining death. In 1968 they cited four criteria as determinative of a permanently nonfunctioning brain: unreceptivity and unresponsivity to externally applied stimuli and inner need; absence of spontaneous movements of breathing; absence of reflexes other than spinal column reflexes, and fixed dilated pupils; and a flat electroencephalogram (EEG) to be used as a confirmatory but not mandatory test.

These tests were to be repeated after 24 h, and in the absence of hypothermia or depressant drugs, were to be considered a confirmation of brain death. It has been conclusively demonstrated that when the flow of blood to the brain has been interrupted for approximately 15 min there is a uniform necrosis and liquefaction of the brain. There has been no documented example in which the Harvard criteria have been shown to be invalid.

Ethical issues. These new criteria for death determination have important consequences of an ethical and social nature; there is concern for the relationship of the criteria to organ transplantation and the role of physicians in establishing the new criteria, and fears concerning future updating of the criteria.

With regard to the issue of organ transplantation, it is contrary to moral intuitions that a dying person might be prematurely defined as dead in order that someone else might receive the organs. Yet, it has been claimed that because anencephalic infants born alive cannot survive because they lack a brain, they should be declared dead and placed on a respirator to maintain circulation, so that any transplantable organs that are needed to sustain the lives of other infants can be removed. However, this is not practical because it is held that a moral obligation exists to treat the living as such until death occurs. But this does not rule out the possibility that a more precise definition of death is needed for the reasons mentioned earlier.

To avoid a conflict in the role of the physician, it has become common practice that the physician for the recipient of an organ donation is not party to the judgment that the donor has died. The needs of a patient for a viable organ must not contaminate the judgment that another person has died. Many states however, have drafted "required request" laws that obligate the hospital to develop a policy to ensure that the opportunity for the family to donate the organs of the newly deceased is not overlooked.

In light of the growing demand for organs suitable for transplant, it has been suggested that the criteria for death be changed. Some physicians and philosophers believe that the Harvard criteria are too stringent. They would, in most cases, favor a definition of cerebral death, a condition in which the neocortex shows a flat (isoelectric) reading in an EEG. Falling into this category would be those who view consciousness and the capacity for social interaction as the essential characteristics of being a living human. It was reported that this criterion for death was used in the former Soviet Union. There is an apparent reluctance to accept this criterion in the West, and the division is largely attributable to the philosophical debate about whether the human being is to be characterized in terms of the biological organism or as a rational and personal being. Thus, there is by no means agreement on this issue, and many physicians and others find it morally abhorrent to consider a person clinically dead according to the neocortex definition when there is still spontaneous breathing and circulation.

Social and legal responses. In addition to the ethical issues stemming from the articulation of a new definition of the criteria for determining death, there have been legal and social implications. Many states, either by legislation or by court opinion, have determined that the irreversible loss of brain function can be used as legal evidence of death.

In 1977 the American Medical Association's House of Delegates adopted the position that "permanent and irreversible cessation of the brain constitutes one of the various criteria which can be used in the medical diagnosis of death."

In August 1978 the National Conference of Commissioners on Uniform State Laws approved a Uniform Brain Death Act and recommended it to the states for adoption. It reads as follows: "For legal and medical purposes, an individual who has sustained irreversible cessation of all functioning of the brain, including the brain stem, is dead. This judgment is to be made in accordance with reasonable medical standards."

Prolonging life. The ability to define death with more precision is important in cases involving the irreversibly comatose individual. It is generally accepted as both a moral and legal right that competent persons have the right to refuse medical treatment. The law of informed consent is based upon this right, and the law has upheld the right of competent adults to decline treatment, even at the cost of their lives. Increasingly, dying individuals claim a more active role in self-determination regarding medical treatment in the final stage of life.

Every state and the District of Columbia have passed laws allowing individuals to either develop a living will or name an agent to hold durable power of attorney for health care decisions at the end of life, should they lose capacity for decision making. The intent is generally to provide the opportunity for individuals to extend their rational choice to refuse certain types of life-prolonging treatment. A signed directive is filed for this purpose and becomes authoritative if the individual is in an unconscious state and death is deemed imminent.

A federal law known as the Patient Self-Determination Act requires all institutions receiving federal funds to notify individuals at the time of

admission to a hospital, hospice, nursing home, or retirement facility of the right to a Natural Death Act Directive (living will), and to designate a person to hold durable power of attorney for health care decisions. While the establishment of new public policy may be helpful in some cases, it does not resolve all of the issues at stake. Many physicians believe a Natural Death Act is helpful, but primarily as an encouragement to opening communication within families. Such a law leaves unanswered the question as to whether persons who fail to sign such a directive beforehand might be at greater risk to have their lives prolonged against their will than before the law. Also, what of special cases such as that of a renal dialysis patient who finds life on the kidney machine intolerable, but for whom death is not imminent if treatment continues?

Even before these legislative innovations, the issue received considerable attention from ethicists and theologians. Pope Pius XII ruled that there was not a moral duty to prolong life in the face of imminent death when the provision of such treatment would be difficult to obtain, dangerous, painful, unusual, or costly and burdensome to the individual or the family. Herein is a moral persuasion, an appeal not to law but to the principles of reasonableness in the light of the prognosis and the relevant human factors in the case. Thus, when the burdens of treatment outweigh the benefits for the individual, it is considered morally correct to stop treatment. Court opinion has often upheld the lawfulness of discontinuing the use of respirators, tube feedings, and hydration.

Euthanasia. Finally, one of the more radical issues is that of actively ending the life of a suffering person or helping another to die through a positive action. Some ethicists see no moral distinction between removing life-support systems or lifesaving medications which allow a patient to die and taking a direct action which causes a patient to die, since the intention and the consequence are the same in both cases. Others disagree, claiming that there is indeed a morally significant difference if a person dies of the underlying disease and is not made to endure still greater suffering by prolonging the dying process than if the person dies because another has caused the death through an intentional and fatal action.

Two events have occurred regarding the controversial issue of physician-assisted suicide. A federal district judge held that the state law making assisted suicide a class C felony was unconstitutional. The ruling is limited to cases involving mentally competent, terminally ill patients who are not under undue influence and who wish to have physician assistance in committing suicide. The second major shift in public policy occurred in Oregon and provides legal protection for physicians who assist dying patients to commit suicide.

Opponents of assisted suicide and active euthanasia are not convinced that removing legal prohibitions are good for society. They prefer to work toward providing suffering individuals with adequate pain control and comfort measures during the last stages of life. They argue that even if in some cases a moral justification could be found for actively ending the life of a suffering dying individual the social consequences that would arise from the general tolerance of such a practice would be more problematic than good.

Thomas R. McCormick

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De Broglie wavelength

The wavelength $\lambda = h/p$ associated with a beam of particles (or with a single particle) of momentum p; $h = 6.63 \times 10^{-34}$ joule-second is Planck's constant. The same formula gives the momentum of an individual photon associated with a light wave of wavelength λ. This formula, along with the profound proposition that all matter has wavelike properties, was first put forth by Louis de Broglie in 1924, and is fundamental to the modern theory of matter and its interaction with electromagnetic radiation. *See* NONRELATIVISTIC QUANTUM THEORY; QUANTUM MECHANICS.

Edward Gerjuoy

Decapoda (Crustacea)

One of the more highly specialized orders of the subphylum (or class) Crustacea. This order includes the shrimps, lobsters, hermit crabs, and true crabs. The order is so diverse that satisfactory definition is difficult, but a few characters are common to nearly all decapods. The most obvious is a head shield, or carapace, which covers and coalesces with all of the thoracic somites and which overhangs the gills on each side. The first three of the eight pairs of thoracic appendages are specialized as maxillipeds and closely associated with the true mouthparts. The last five pairs are walking legs or chelae (claws); this feature is the basis of the name Decapoda (= ten feet). The gills are usually well developed and arranged in several series.

Adult decapods vary in size from less than 1/2 in. (13 mm) in length to that of the giant Japanese crab (*Macrocheira*), the largest living arthropod, a spider crab which may span more than 12 ft (3.6 m) between the tips of the outstretched claws. Although most decapods are found in the sea, they are by no means restricted to that habitat. Crayfishes are well-known inhabitants of freshwater streams and ponds, as are several kinds of shrimps and some true crabs. A number of crabs and hermit crabs have become well adapted to a terrestrial existence far from water; they return to the sea only seasonally to hatch their eggs. Many crayfishes burrow in the ground, and one

species of crab spends its entire life in the tops of lofty trees.

In a general way, the decapods may be divided into two groups: the long-tailed and the short-tailed forms. The long-tailed species, such as the shrimps and lobsters, have a more or less cylindrical, or laterally compressed, carapace that often bears a head spine or rostrum. The large, muscular abdomen permits the shrimp or lobster to dart quickly backward out of danger, but the succulence of the tail makes the animals prey to numerous enemies, including humans. In the short-tailed species, the crabs, the carapace is often broadened and flattened dorsally and usually does not form a rostrum. The much-reduced and feeble abdomen is tucked under the thorax, where it serves the female as a brood pouch for the eggs.

Appendages

Like other malacostracans, nearly all decapods have pairs of appendages on each of 19 segments: 5 pairs on the head, 8 on the thorax, 6 on the abdomen (described in sequence from anterior to posterior ends). Each of the first pair of antennae (also called antennules) has a three-segmented peduncle and, typically, two flagella, the outer one being bifurcated near the base in some shrimps. The second antennae have up to five segments in the peduncle, a single flagellum, and often a scalelike outer branch or exopod (**Fig. 1**).

Mouth. There are six pairs of mouthparts in the decapods, the posterior three pairs being modified thoracic legs. The mandibles sometimes have distinct incisor and molar processes. These are sometimes indistinguishably fused, or the incisor process may be lacking. There is usually a three-segmented palp (**Fig. 2**). The first maxillae (maxillulae), second maxillae (maxillae), and the first maxillipeds are usually foliaceous. The second maxillipeds are also composed of broad, flattened segments, but they are less modified from the typical thoracic appendages than the first maxillipeds. The third maxillipeds are elongate and leglike in the shrimps but are greatly modified, forming an operculum over the preceding mouthparts in most of the crabs.

Thorax. The five remaining pairs of thoracic appendages, the pereiopods (or pereopods), are basically walking legs although one or more pairs may be variously modified. The most common modification is seen in the formation of pinching chelae. In the long-tailed decapods, one to three (rarely four or all five) pairs of pereiopods may be chelate. In the short-tailed forms, only the first pair is usually thus modified. Some of the pereiopods may be broadened and flattened as swimming paddles, especially in the swimming crabs. In the shrimps, the pereiopods are basically composed of seven segments, of which some are occasionally multiarticulate. In most of the other decapods, however, the second and third segments are fused. Occasionally, one or two pairs of pereiopods are vestigial or entirely lacking.

Abdomen. The first five pairs of abdominal appendages, the pleopods, are swimming organs in

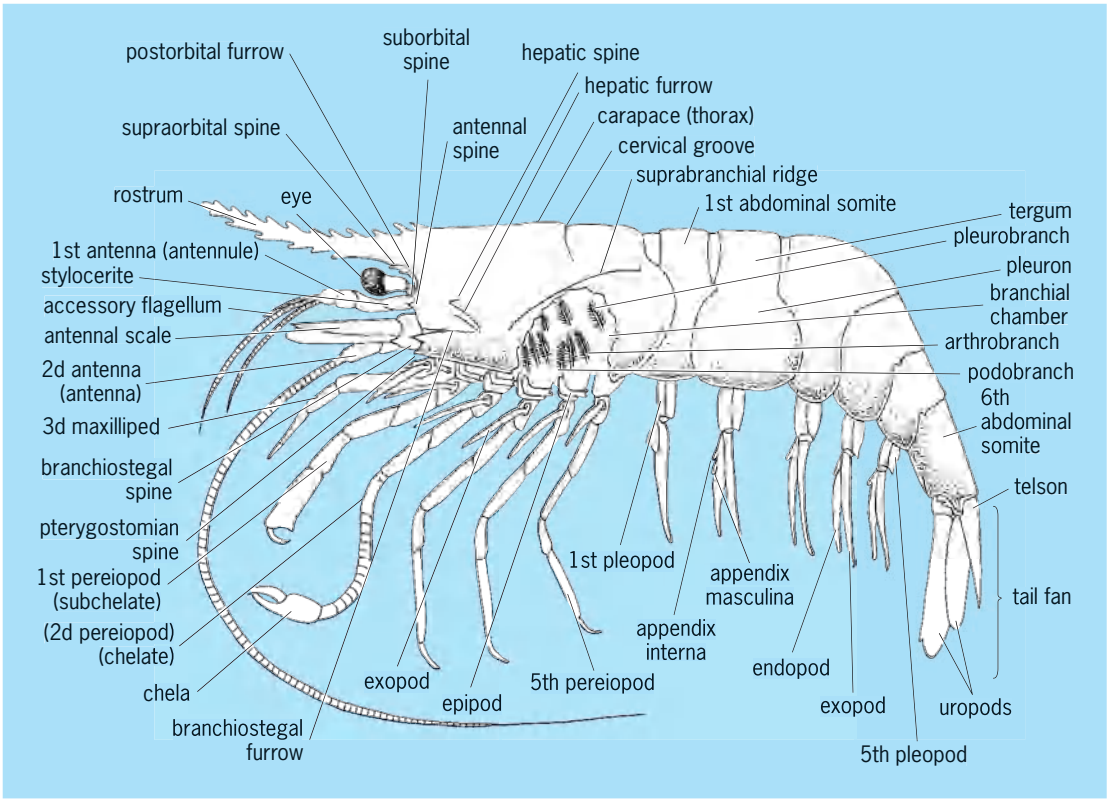

Fig. 1. Lateral view of caridean shrimp showing external morphology.

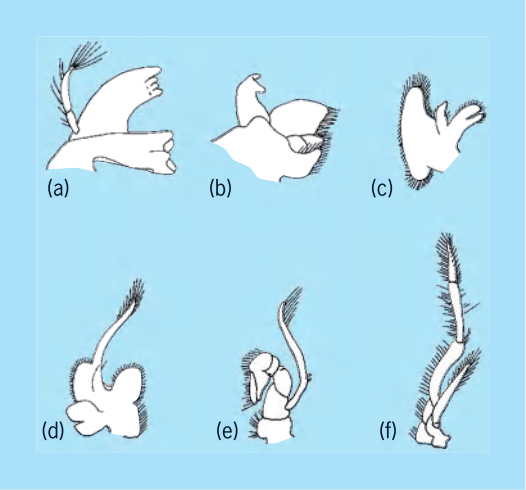

Fig. 2. Mouthparts of caridean shrimp, *Palaemon*. (*a*) Mandible showing incisor and molar processes and palp. (*b*) First maxilla (maxillule). (*c*) Second maxilla (maxilla). (*d*) First maxilliped. (*e*) Second maxilliped. (*f*) Third maxilliped.

most of the long-tailed decapods, but the first one or two pairs are usually modified as sexual appendages (gonopods) in the male and are often reduced and sometimes absent in the female. In the short-tailed forms, the pleopods (except those in the male which are modified as male sexual organs) are usually reduced or modified in the female for the attachment of eggs. In both sexes of the long-tailed species, the last pair of abdominal appendages, the uropods, forms a tail fan with a terminal median structure, the telson. In the hermit crabs, the uropods are thickened and twisted to hold the animal in the gastropod shell or other structure chosen as a shelter. In practically all of the crabs, the uropods are absent.

Nervous System and Sense Organs

The nervous system is somewhat variable. Although usually fewer, there may be as many as 11 ganglia, 5 thoracic and 6 abdominal, in the ventral nerve chain posterior to the subesophageal ganglion (**Fig. 3**). In most of the crabs, all of the ventral ganglia are concentrated in a rounded mass in the thorax, from which the nerves radiate outward.

The sense organs include a pair of compound eyes and statocysts, olfactory filaments, and tactile setae. The eyes are well developed in most decapods, but they may be reduced or entirely absent in species living in the deep sea, in caves, or in burrows. They are borne on limblike eyestalks located near the antennae and composed of two, rarely three, segments, either of which may be abnormally elongate. The cornea is usually distinctly faceted, the facets being square or hexagonal. The statocysts, or balancing organs, occur on the first segment of the antennules. They are cavities, either open or closed to the exterior, lined with sensory hairs. If open, they usually contain foreign particles, such as sand grains. If closed, they may contain secreted bodies, but usually there is no inclusion. Olfactory filaments are especially numerous on the outer flagellum of the antennules. Tactile setae, or hairs, occur on various parts of the body and appendages.

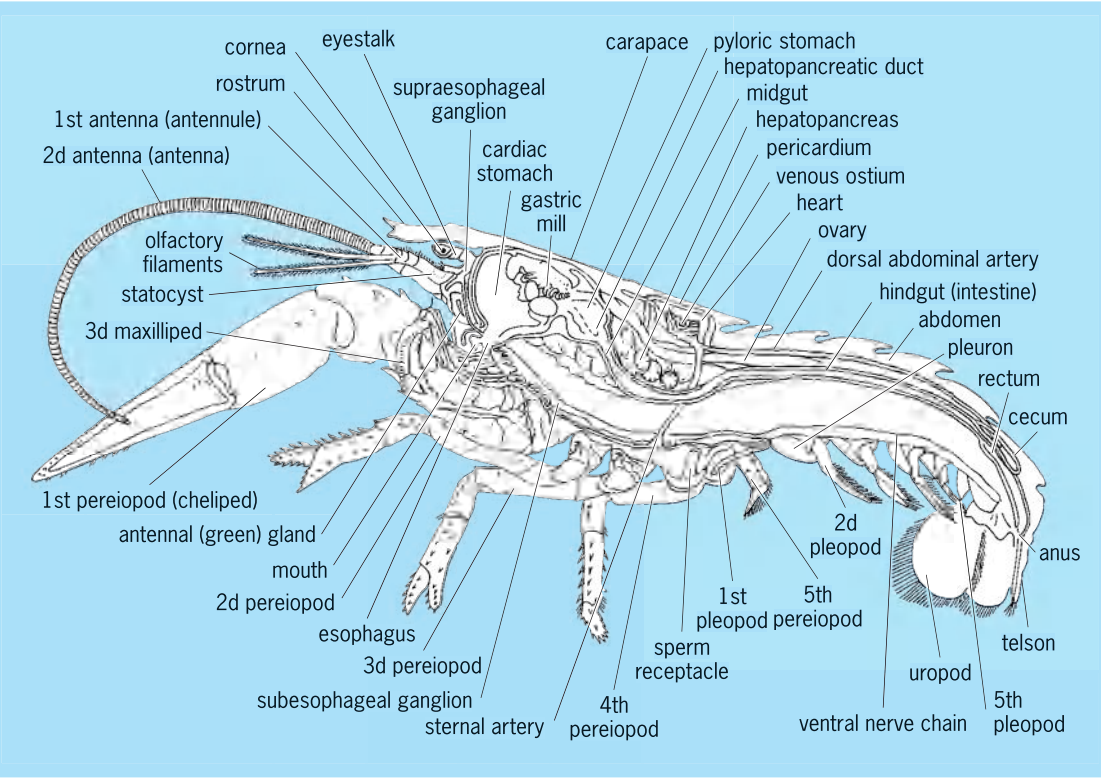

Fig. 3. Median section of lobster, *Homarus*, showing general anatomy.

Alimentary and Circulatory System

As in most crustaceans, the alimentary tract is a more or less straight tube divided into three parts: the foregut or stomodeum, the midgut or mesenteron, and the hindgut or proctodeum. The first and last have a chitinous, and sometimes calcified, lining. The foregut, in most decapods, is dilated to form a stomach. Its anterior, or cardiac, portion is at least partially lined with movably articulated, calcified ossicles, controlled by a complex system of muscles called the gastric mill. The posterior, or pyloric, part is provided with hairy ridges which combine to form a straining apparatus. The midgut is variously furnished with blind tubes, and a mass of minutely ramified tubules, known as the hepatopancreas or liver, empties through one or more ducts into the anterior end of the midgut.

The heart is short and polygonal, and situated under the posterior part of the carapace. It has three or more pairs of ostia (holes with valves) and two pairs of arteries lead from it, in addition to the median anterior and posterior arteries. The blood collects in a venous sinus in the midventral line, passes to the gills, and then back to the pericardial sinus surrounding the heart.

Respiration

In practically all decapods, respiration is accomplished by a series of gills, attached to the lateral walls of the thoracic somites and to the basal segments of the thoracic appendages. A complete set of gills consists of the pleurobranchs, attached to the lateral walls of the thoracic somites; the arthrobranchs, fastened to the articular membranes between the body wall and the first segment of the appendage; and podobranchs, attached to the first segment of the appendage. The gills (**Fig. 4**) may be one of three types: trichobranchiate, with filamentous branches arranged in several series around the axis; phyllobranchiate, with flattened branches usually arranged in two opposite series; and dendrobranchiate, with extensive, sometimes complex, branching of the two primary series.

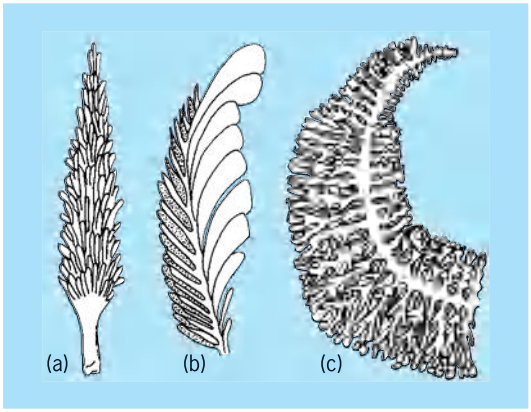

Fig. 4. Decapod gills. (*a*) Trichobranchiate gill of stenopodid shrimp, *Spongicola*. (*b*) Phyllobranchiate gill of mud shrimp, *Thalassina*. (*c*) Dendrobranchiate gill of penaeid shrimp, *Benthesicymus*.

A continuous current of water is drawn through the branchial chamber and over the gills by the fluttering motion of the outer lobe (scaphognathite) of the second maxilla. The current usually moves forward, but this direction may be periodically reversed in burrowing species. In the latter, also, the exhalant channels may be extended anteriorly by the apposition of the left and right elements of certain appendages, such as the first and second antennal flagella, the antennal scales, the mandibular palps, and the third maxillipeds.

In those crabs best adapted for life on land, the action of the gills is supplemented by the lining membrane of the branchial chamber. This chamber is covered with minute villi and is unusually well supplied with blood vessels, and thus functions as a lung. A similar highly vascularized area occurs in the delicate abdominal skin of the terrestrial hermit crabs.

Excretion

The antennal or "green" gland is the chief excretory organ in all decapods. Maxillary glands are often present in the larval stages, but they never persist in the adult. The antennal gland is compact and more complex than in any other crustacean group. It is often divided into regions: the saccule (end sac), which is usually partitioned or ramose; the labyrinth, a spongy mass with a complex system of folds; and a coiled canal and bladder with a duct leading to the exterior (**Fig. 5**). The bladder may be a simple vesicle

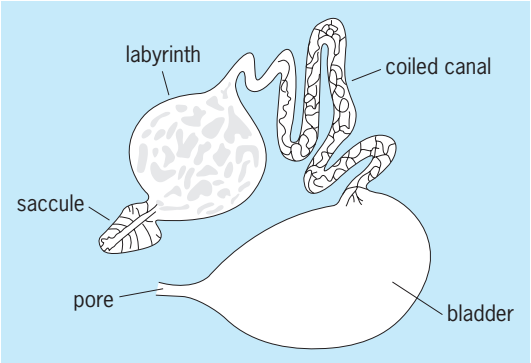

Fig. 5. Antennal gland of crayfish (section).

with a short duct, as in the crayfish; or it may be variously lobate, as in most of the shrimps, crabs, and hermit crabs. In the latter case, two diverticula from the bladder extend through the entire abdomen. The external opening is usually in a small elevation on the first segment of the second antennae. In the crabs, the aperture is covered by an operculum which may be opened and closed.

Reproductive System

The sexes are separate in most decapods, although protandric hermaphroditism (in which the animal develops first as a male and subsequently changes to a female) occurs in some shrimps. The testes most often lie partly in the thorax and partly in the abdomen, and they are usually connected across the

midline. In some hermit crabs, however, they lie wholly within the abdomen, and they may be either unconnected or fused into a single organ. The vas deferens usually opens on the first (most proximal) segment of the last pair of thoracic legs, but in some crabs it perforates the last thoracic sternum. The ovaries resemble the testes in shape and position. The oviducts usually open on the first segment of the third pereiopods; in most crabs, however, they open on the sternum.

Development

In only one group of shrimps (suborder Dendrobranchiata) are all of the presumably primitive larval stages represented (**Fig. 6**), some with multiple instars separated by molts. The first stage is a typical crustacean nauplius with an unprotected oval body bearing a median simple eye and three pairs of appendages that are used in swimming. This stage is followed by a metanauplius, in which four more pairs of appendage rudiments appear. The third pair of the original appendages becomes less of a swimming organ and more like a pair of primitive mandibles. In the third stage, the protozoea, the seven pairs of appendages become more highly developed, a carapace covers the anterior part of the body, the abdomen is clearly formed though unsegmented at first, the rudiments of paired eyes appear, and the heart is formed. In the next stage, the zoea, the eyes become movable and the carapace develops a rostrum. All thoracic appendages are present, at least in rudimentary form, and those of the abdomen appear, especially the uropods. In later zoeal instars (mysis stage), the well-developed thoracic

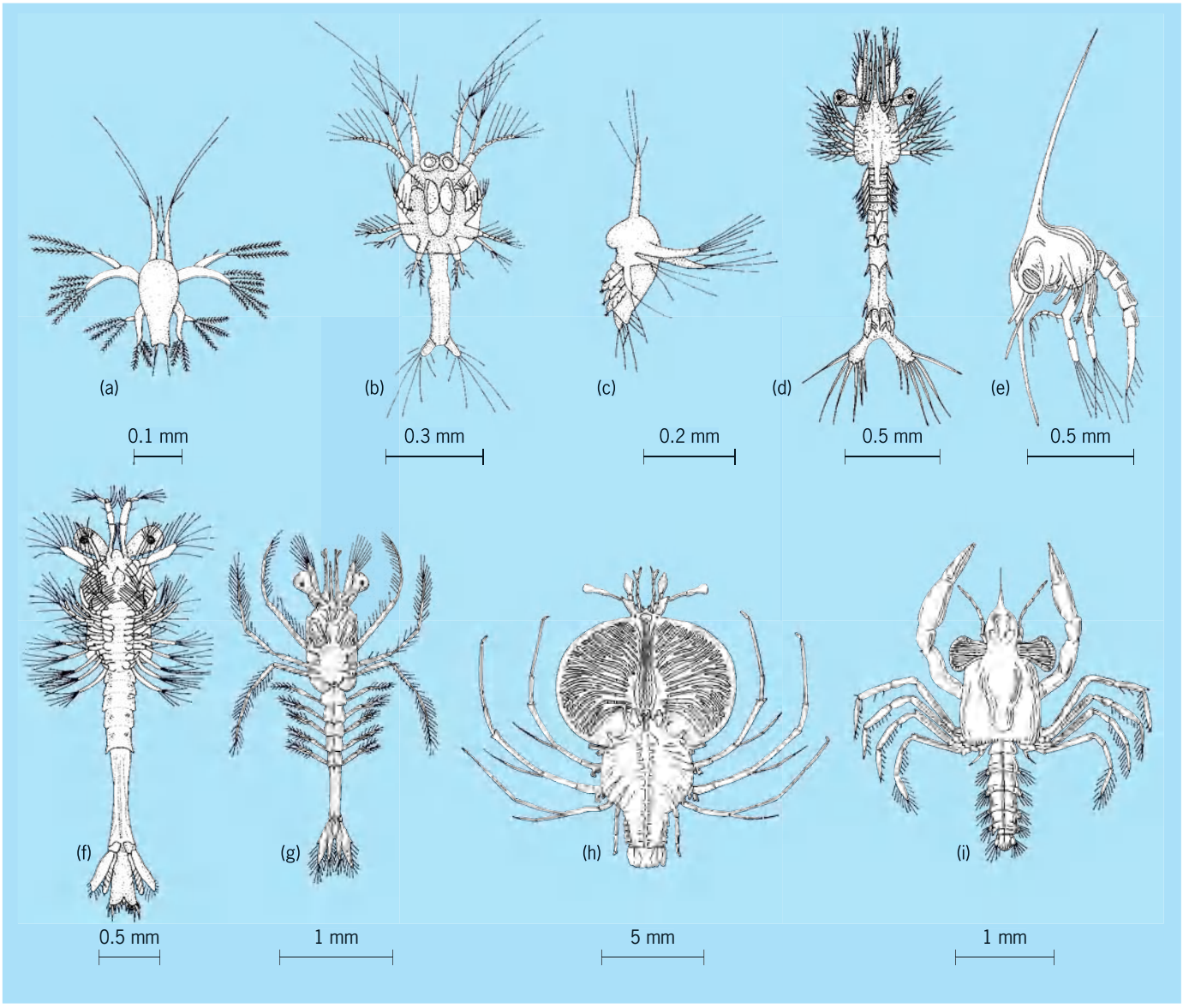

Fig. 6. Decapod larval stages. (*a*) Nauplius larva of penaeid shrimp (dorsal view). (*b*) Protozoea larva of penaeid shrimp. (*c*) Metanauplius larva of penaeid shrimp (lateral view). (*d*) Metazoea larva of penaeid shrimp. (*e*) Zoea larva of crab. (*f*) Mysis larva of penaeid shrimp. (*g*) Mysis (acanthosoma) larva of sergestid shrimp. (*h*) Phyllosoma larva of scyllarid lobster. (*i*) Megalopa larva of swimming crab.

appendages replace the antennae as the chief swimming organs. The abdomen increases markedly in size and takes on a form similar to that of the adult.

In most decapods (constituting the suborder Pleocyemata), the nauplius stage is passed while in the egg, and the animals hatch as zoeae, often molting several times as they differentiate and grow. The last zoeal instar metamorphoses into a postlarva or megalopa. This is similar to the adult, except that all appendages are present, the abdomen is prominent, and the pleopods are used for swimming.

Decapods that live in deep-water or polar habitats tend to exhibit a reduction in the number of larval instars; in freshwater species there may be none. Most freshwater shrimps hatch in a late zoeal (mysis) stage, and the crayfishes and freshwater crabs hatch as miniature adults. Terrestrial forms, on the other hand, typically migrate to the sea to hatch their eggs, which then pass through the usual larval stages of marine species.

Phylogeny

No true decapod fossils are known with certainty from Paleozoic deposits, but decapod shrimps are not uncommon in Triassic and, especially, Jurassic rocks. Lobsterlike forms related to Recent deep-sea species also appeared in the Triassic; true lobsters are known from the Jurassic, and crayfishes from the Cretaceous. Crabs are well known as fossils, with primitive groups appearing in the Jurassic; several more specialized groups are found in the Cretaceous, and all of the major existing groups were represented in the Tertiary. *See* CRETACEOUS; JURASSIC; TERTIARY.

The evolution of the decapods has not been satisfactorily worked out and may never be known completely, but several hypotheses based on cladistic analysis of morphology and deoxyribonucleic acid (DNA) sequences have been put forth recently. The presence of a nauplius larva in the development of the dendrobranchiate shrimps suggests a link with the more primitive malacostracan crustaceans, such as the Euphausiacea (krill), but it can hardly be doubted that the Decapoda, as a group, are highly specialized.

Classification

The classification of the approximately 18,000 living species of decapods presents a difficult problem that has not yet been entirely solved. The division of the order into long-tailed (suborder Natantia) and short-tailed (suborder Reptantia) groups, alluded to previously, was in favor in the eighteenth century. This scheme was generally abandoned, however, when it was discovered that some of the long-tailed forms (known in older classifications as section Macrura, including the lobsters and crayfish) were obviously more closely related to the short-tailed group than they were to other long-tailed species. The chief disparity seems to lie between the shrimps that possess dendrobranchiate gills and all of the other decapods (Pleocyemata), in which the gills are either trichobranchiate or phyllobranchiate. The current, generally accepted classification scheme, follows.

Order Decapoda
- Suborder Dendrobranchiata
- Suborder Pleocyemata
  - Infraorder: Caridea
    - Stenopodidea
    - Palinura
    - Astacidea
    - Thalassinidea
    - Anomura
    - Brachyura

Dendrobranchiata. This suborder (equivalent to the section Penaeidea of older classifications) contains the superfamilies Penaeoidea and Sergestoidea, which make up most of the shrimps caught or "farmed" worldwide (**Fig. 7**). Collectively they are distinguished by the following characteristics: the pleura (skeletal plates along the sides) of the first abdominal somite overlap those of the second; the third legs are nearly always chelate but are no stouter than the first pair; the first pleopods of the male bear a complicated flaplike appendix, the petasma, while the females have a spermatophore receptacle, the thelycum, on the ventral surface of the posterior thoracic somites. Unlike those of all other decapods, the

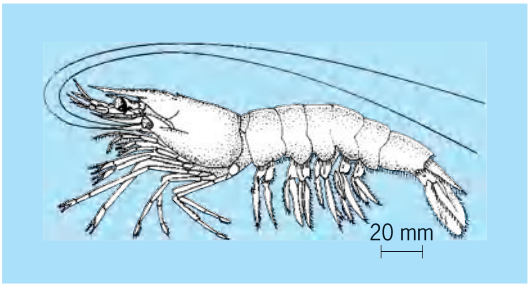

Fig. 7. Penaeidean shrimp, *Litopenaeus setiferus*, a primitive decapod.

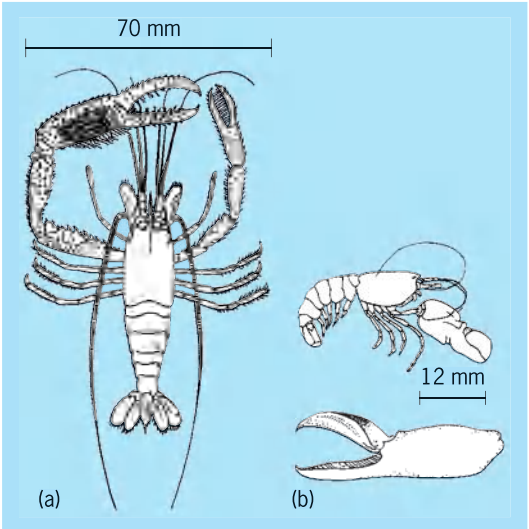

Fig. 8. Caridean shrimp. (*a*) River shrimp, *Macrobrachium faustinum*. (*b*) Snapping shrimp, *Alpheus heterochaelis*, with minor claw enlarged.

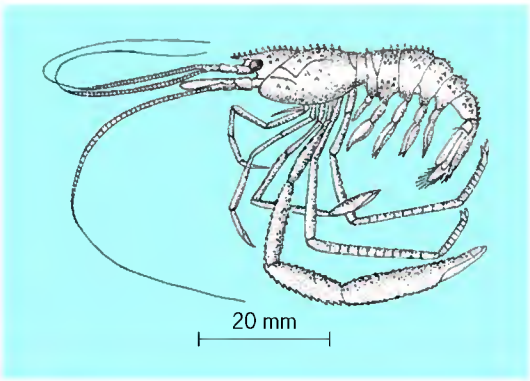

Fig. 9. Stenopodid shrimp, *Stenopus hispidus*, a marine decapod.

eggs are not attached to the abdominal appendages but are almost always shed free into the water. The Penaeoidea share with the Euphausiacea the presence of a petasma; in both groups, the eggs are rarely carried, and they hatch at an early larval stage, usually as nauplii. The best-known penaeoideans, including most of the commercially important shrimps or prawns of the warmer seas (family Penaeidae), live on muddy bottoms in shallow or moderate depths. However, they also occur both in the deep sea and as pelagic organisms in the middepths (around 500 m or 1640 ft). In this latter region are found the Sergestidae, in which the last two pairs of thoracic legs are reduced or absent; some have luminescent organs (photophores). There are

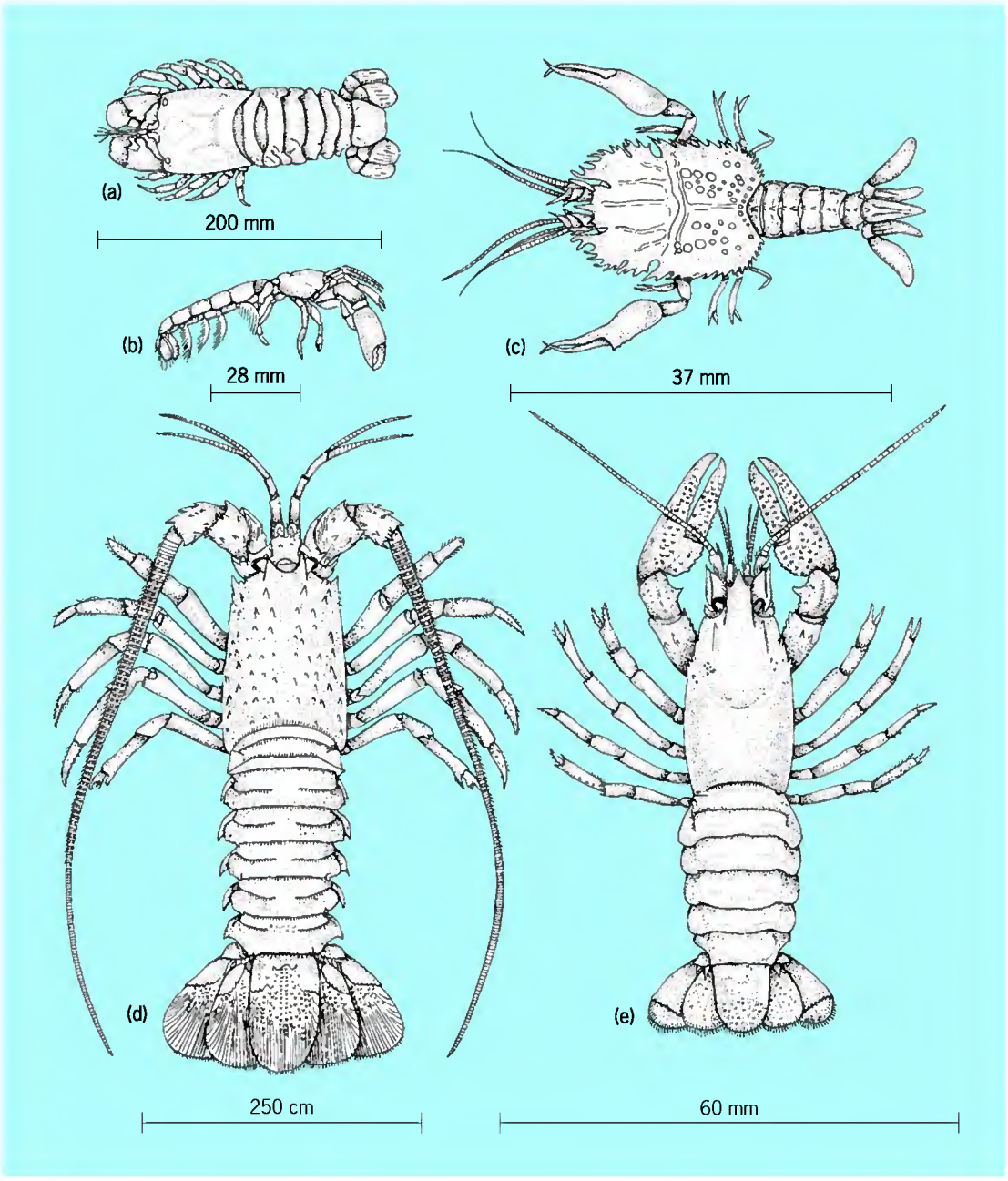

Fig. 10. Macruran decapods. (*a*) Spanish lobster, *Scyllarides aequinoctialis*. (*b*) Mud shrimp, *Callianidea laevicauda*. (*c*) Eryonid, *Polycheles crucifer*. (*d*) Spiny lobster, *Panulirus interruptus*. (*e*) Crayfish, *Orconectes limosus*.

about 500 species of living Dendrobranchiata.

Pleocyemata. This suborder contains the infraorders Caridea, Stenopodidea, Palinura, Astacidea, Thalassinidea, Anomura, and Brachyura.

Caridea. This is the largest and most diverse group of shrimps and prawns. In this group, the pleura of the second abdominal somite overlap those of the first; the third legs are never chelate; there is no petasma or thelycum; and the gills are phyllobranchiate. The sexes can usually be distinguished by the presence in the male of two small stalks, an appendix masculina and an appendix interna, located on the inner edge of the inner branch (endopod) of the second pleopods; only the appendix interna occurs in the female. Carideans are found in all parts of the sea, often in association with other marine animals. Members of at least two families, the Atyidae and Palaemonidae, are also widespread in freshwater. The edible shrimps and prawns of northern Europe and North America, of the genera *Crangon* and *Pandalus*, belong to the Caridea, as do the large freshwater prawns of the tropical genus *Macrobrachium*. Some of the latter are the largest natantians known, attaining a length of more than a foot and having abnormally long second pereiopods, which when fully grown may be as long as the body (**Fig. 8***a*). The snapping shrimps, or Alpheidae, make a loud popping noise by cavitation, using a socket-and-plunger mechanism on an enlarged first pereiopod (claw) [Fig. 8*b*]. The resulting crackling noise from vast populations in tropic seas sometimes interferes with underwater sound equipment. Most shrimps and prawns are favorite foods of fishes, but some carideans maintain "service stations" for fishes where they remove parasites from the outer skin, mouths, and gills of the fish. More than 2800 species of Caridea are known.

Stenopodidea. This is a small group of shrimps which superficially resemble the Penaeoidea (**Fig. 9**). The third pereiopods are chelate but are much longer and stouter than the first pair; the pleura of the second abdominal somite do not overlap those of the first; there is no petasma or thelycum; and the gills are trichobranchiate. All of the 50 or more living species are marine; some are closely associated with sponges and corals.

Palinura, Astacidea, and Thalassinidea. The infraorders Palinura, Astacidea, and Thalassinidea, constituting the Macrura of older classifications, are characterized by an extended abdomen, well-developed tail fan, and trichobranchiate gills (**Fig. 10**). They are capable of swimming, as shrimps do, by flexion of abdominal muscles. However, like crabs, they are better adapted for crawling over the substrate, with strong legs, or burrowing into it.

Palinura is a traditional infraorder (which is probably not monophyletic) containing two major groups (superfamilies), Palinuroidea and Eryonoidea, and about 200 species. The Palinuroidea (Achelata in some contemporary classifications) include the spiny lobsters or langoustes (family Palinuridae) [Fig. 10*d*] and the Spanish, or shovel-nosed, lobsters (Scyllaridae) [Fig. 10*a*]. They are heavily armored like the true lobsters but are distinguished by the absence of a rostrum and of chelae, except occasionally on the last pereiopod of the female. They are abundant in shallow and moderate depths of warm and temperate seas, where they are often of considerable commercial importance. Frozen tails of spiny lobsters are imported to the United States from South Africa, Australia, and New Zealand, and the fresh lobster meat is the basis of fisheries in the Caribbean, southern Europe, and Japan. The Eryonoidea (Polychelida in some recent classifications) are rather thin-shelled, blind inhabitants of the depths of the sea (Fig. 10*c*). The carapace is flattened dorsally and considerably expanded laterally. The first four, or all five, pairs of pereiopods are chelate, with the first pair much elongated. This group was probably more numerous in ancient seas.

Astacidea are the true lobsters and crayfishes, characterized by a rostrum and by chelae on the first three pairs of pereiopods, with the first pair being noticeably larger than the others (Fig. 10*e*). Most lobsters (Nephropidae) are found in cool seas or in the cool, offshore waters of the tropics. The most familiar lobsters (*Homarus*) are those found along the Atlantic coasts of Europe and North America. The Norway lobster of Europe (*Nephrops*) is also of some commercial importance but is smaller and less meaty. The crayfishes (Cambaridae, Astacidae, Parastacidae) are widespread in freshwaters of the temperate regions of all continents except Africa. They are of commercial importance in southern Europe and Australia. More than 700 living species of lobsters

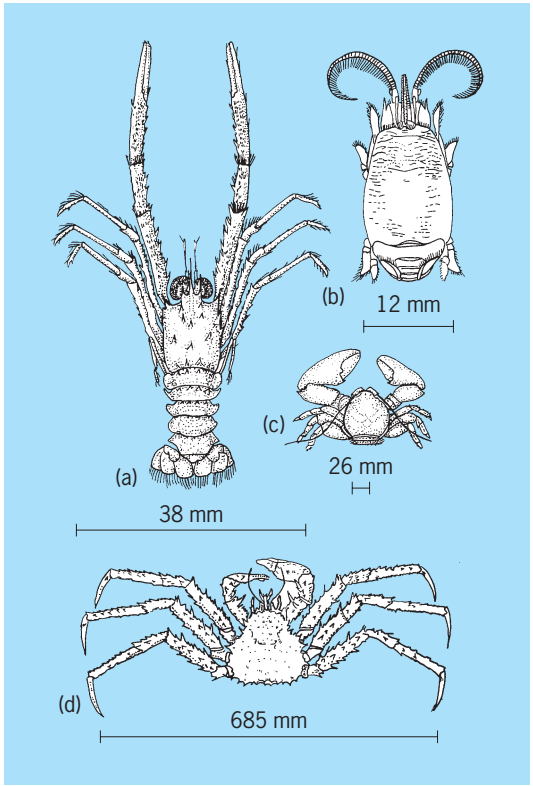

Fig. 11. Anomuran decapods. (*a*) Galatheid, *Munida evermanni*. (*b*) Mole crab, *Emerita talpoida*. (*c*) Porcellanid crab, *Petrolisthes tridentatus*. (*d*) King crab, *Lithodes maia*.

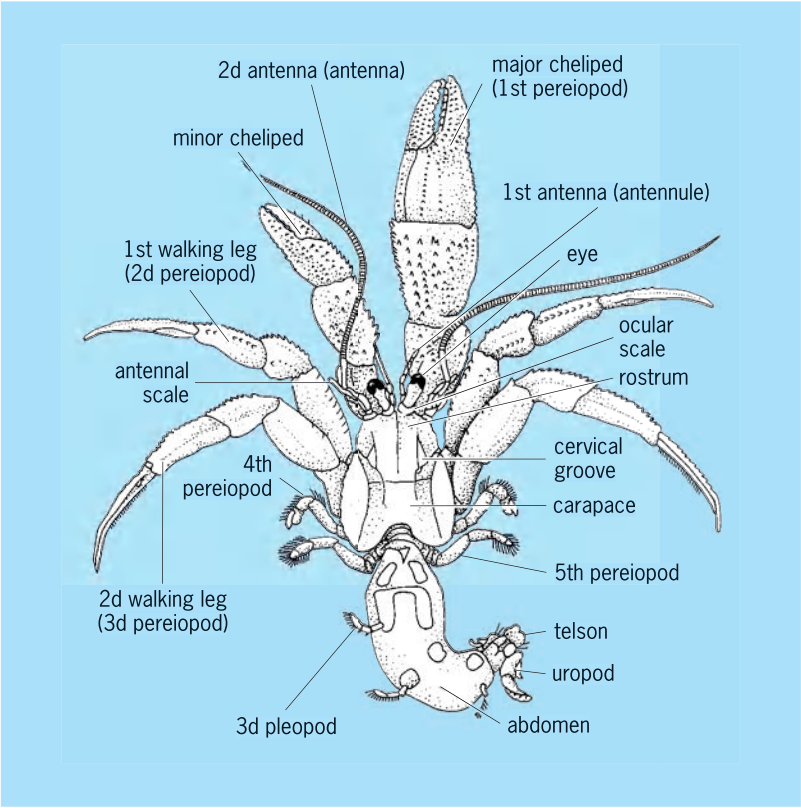

Fig. 12. Dorsal view of hermit crab, *Pagurus*, showing external morphology.

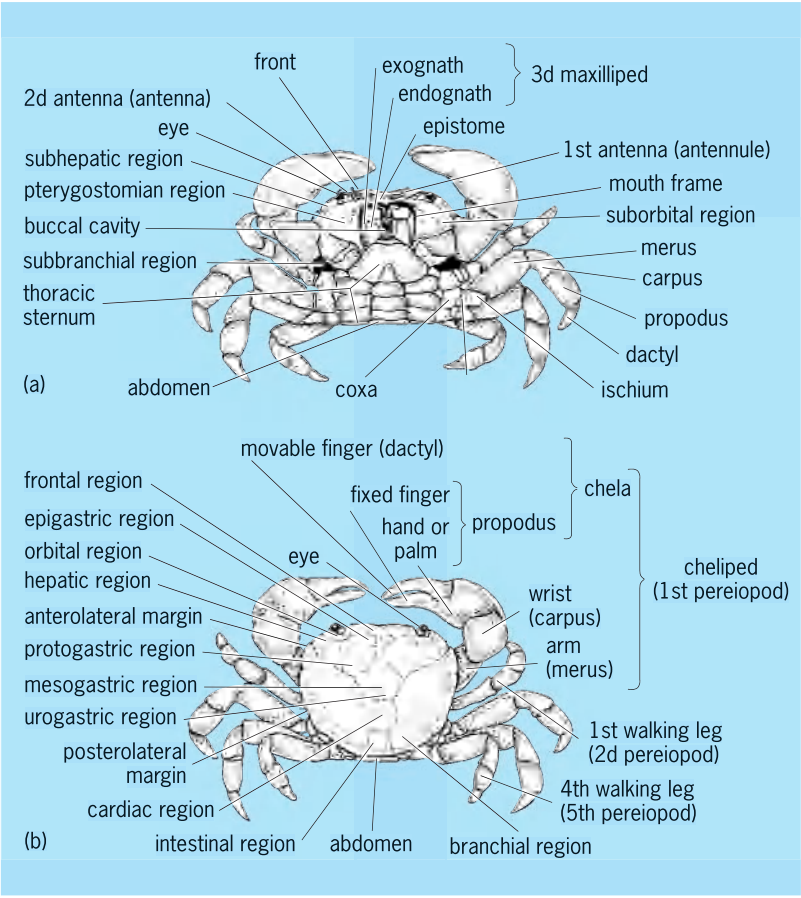

Fig. 13. External morphology of grapsoid crab. (*a*) Ventral view. (*b*) Dorsal view.

and crayfishes are known, more than half of them being from the freshwaters of the United States.

Thalassinidea are usually thin-shelled, burrowing crustaceans with large, chelate or subchelate first pereiopods, and no chelae on the third pereiopods (Fig. 10*b*). Known as mud or ghost shrimps (Upogebiidae and Callianassidae, respectively), the more than 500 living species are found in shallow and deep seas throughout the world, especially in tropical and warm temperate regions.

Anomura. This possibly unnatural group (renamed Anomala in some newer classifications) contains about 1900 species that are morphologically intermediate between the Macrura and the Brachyura (**Fig. 11**). It includes the galatheids and porcelain crabs (superfamily Galatheoidea), the hermit and king crabs (Paguroidea), and the mole crabs or hippas (Hippoidea). In nearly all of these diverse crustaceans, the first, and sometimes the last, pereiopods are chelate or subchelate. The abdomen is usually bent forward ventrally or is asymmetrical, soft, and twisted. Most anomurans possess phyllobranchiate gills.

The Galatheoidea include those anomurans with a symmetrical abdomen bent upon itself and provided with a well-developed tail fan. Most species are found in the sea at shallow or considerable depths, but one aberrant genus (*Aegla*) inhabits freshwater streams of temperate South America. The more primitive species (Galatheidae) are somewhat lobsterlike (Fig. 11*a*); the porcelain crabs or rock sliders (Porcellanidae) resemble the true crabs but may be distinguished by the much reduced and chelate fifth pereiopods and by the well-developed tail fan (Fig. 11*c*). They are most abundant in the intertidal zone and in moderate depths of the warmer seas; some species live in symbiotic relationships with other animals such as tube-dwelling polychaete worms and sponges.

In the hermit and king crabs (Paguroidea), the abdomen is nearly always asymmetrical, being either soft and twisted or bent under the thorax. The uropods, when present, as in the hermit crabs, are adapted for holding the body in a portable shelter. The first, and often the last, pereiopods are chelate. The hermit crabs (Paguridae and Diogenidae; **Fig. 12**) are found everywhere in the sea and are especially abundant in the intertidal zone. A few species (Coenobitidae) are also found on land in the tropics. Most hermit crabs live in the shells of dead gastropod mollusks which they must replace periodically as they grow. They sometimes place a sea anemone on the shell. The king crabs (Lithodidae) [Fig. 11*d*] are usually large crablike anomurans which live in offshore marine waters. They can be distinguished from the spider crabs, which they superficially resemble, by the reduced last pereiopods and by the asymmetrical arrangement of the abdominal plates in the female. The king crab of the North Pacific (*Paralithodes*) is of considerable commercial importance as food.

The third group of the Anomura, the Hippoidea (Fig. 11*b*), includes cylindrical or squarish

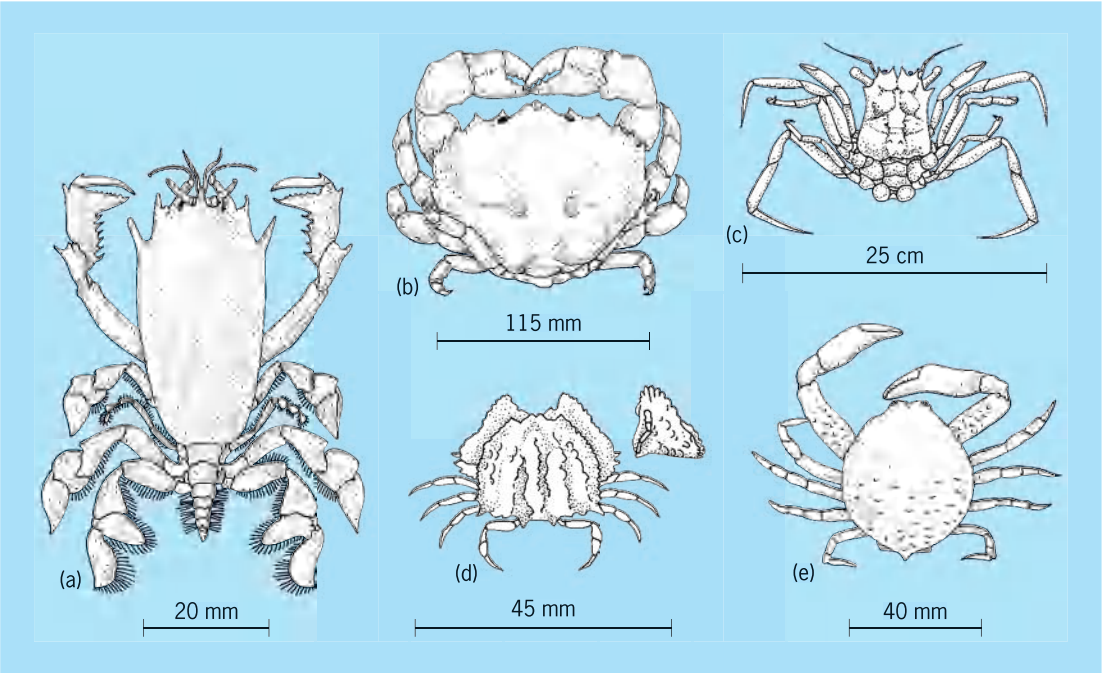

Fig. 14. Common brachyuran crabs. (*a*) Raninoid crab, *Raninoides louisianensis*. (*b*) Dromiid crab, *Dromia erythropus*. (*c*) Mask crab, *Ethusa mascarone americana*. (*d*) Box crab, *Calappa sulcata*. (*e*) Purse crab, *Persephona punctata*.

burrowing crustaceans in which the abdomen is symmetrical and bent under the thorax; the tail fan, if present, is not adapted for swimming; and the first pereiopods are either styliform or subchelate. Common hippas live in sand in the surf zone of tropical and temperate shores, and they travel up and down the beach with the tide. Using their flattened legs, they quickly burrow their entire body into the sand except for their feathery antennae, which strain food particles as the water of receding waves washes over them.

Brachyura. These are the true crabs (**Figs. 13**, **14**, and **15**), with dorsoventrally flattened and widened bodies. The abdomen is symmetrical, without a tail fan, and bent under the thorax. The first pereiopods are always chelate or subchelate, and the gills are phyllobranchiate. The true crabs are as numerous as all other decapods combined, numbering more than 10,000 species. They are divided into about 70 families, grouped principally on the basis of the location of the male and female gonopores (openings of the reproductive system).

Two primitive groups (collectively known as the Podotremata, with gonopores at pereiopod bases) are the Raninoidea (equivalent to the Gymnopleura of older classifications) and the Dromiacea. The carapaces of raninoids are more or less trapezoidal or elongate, the first pereiopods are subchelate, and some or all of the remaining pereiopods are flattened and expanded for burrowing (Fig. 14*a*). In dromiid crabs (Fig. 14*b*), the first pereiopods are chelate; the last pair is dorsal in position and modified for holding objects (such as sponges, tunicates, and bivalve mollusk shells) over the crab. The mouth frame is quadrate.

Most families of crabs are contained within the subsection Heterotremata, defined by the presence of female gonopores on the thoracic sternum rather than on the legs (as in males). In three relatively primitive families, collectively known as the Oxystomata in older classifications, the first pair of pereiopods are chelate and the last pair are either normal or modified as in the Dromiacea. The mouth frame is triangular and forward, over the epistome. The oxystomes include the mask crabs (Dorippidae), box crabs (Calappidae), and purse crabs (Leucosiidae) [Fig. 14*c*, *d*, *e*].

In the remaining heterotremate families (including, in part, the Oxyrhyncha and Brachygnatha of older classifications), the mouth frame is quadrate and the last pereiopods are rarely reduced or dorsal in position. To this group belong the spider, or decorator, crabs (Majidae) [Fig. 15*a*], slow-moving animals that often conceal themselves by attaching seaweeds and various sessile animals to their carapaces. Some cancroid crabs (for example, *Cancer*) are large and abundant enough to be of commercial importance in northern Europe and North America. The rock, or Jonah, crabs of New England and the Dungeness crab of the Pacific Coast are familiar edible crabs of this genus. In the swimming crabs (Portunidae) [Fig. 15*f*], the last pereiopods are modified as swimming paddles; a well-known example is the edible blue crab (*Callinectes*) of the shallow waters of the tropical and temperate Americas. The ubiquitous little mud crabs (Xanthidae) [Fig. 15*e*], well known to every visitor to rocky shores, are characterized by strong, often black-tipped claws. The stone crab (*Menippe*), another xanthoid (Fig. 15*g*), is highly esteemed by connoisseurs of seafood. Also worthy of mention are freshwater crabs (Potamidae, Pseudothelphusidae, and

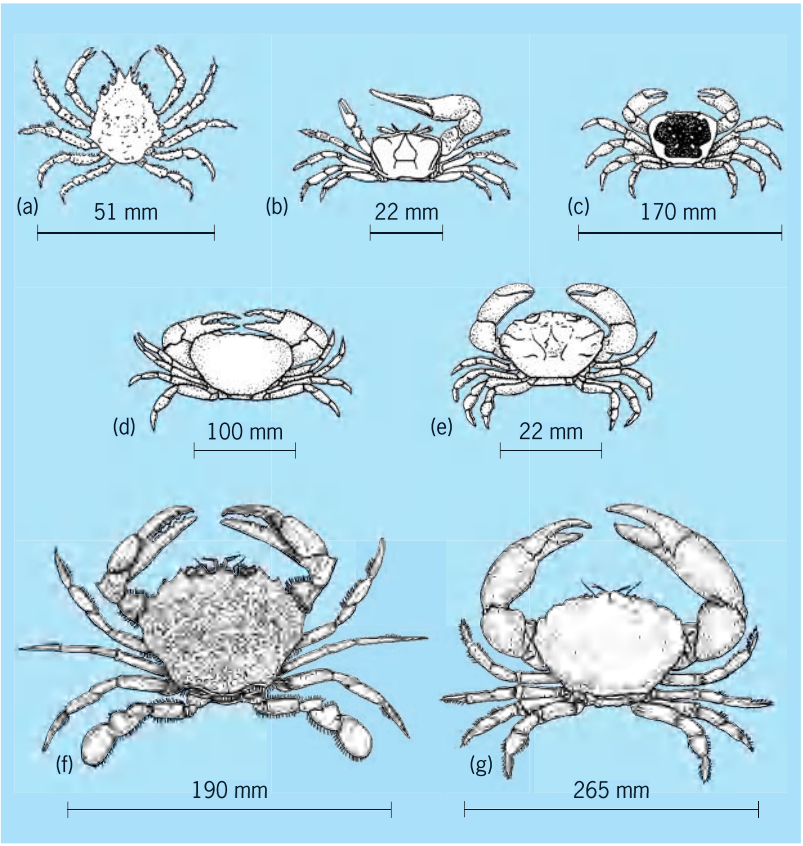

Fig. 15. Other common brachyuran crabs. (*a*) Spider crab, *Mithrax acuticornis*. (*b*) Fiddler crab, *Uca pugilator*. (c) Land crab, *Gecarcinus lateralis*. (*d*) Freshwater crab, *Epilobocera sinuatifrons*. (e) Mud crab, *Eurypanopeus abbreviatus*. (*f*) Swimming crab, *Ovalipes ocellatus*. (*g*) Stone crab, *Menippe mercenaria*.

Trichodactylidae) [Fig. 15*d*], found in tropical and some temperate regions, usually in areas not inhabited by crayfishes.

The last few families are assigned to the subsection Thoracotremata. In this group, the gonopore is located on the sternum in both sexes. The Pinnotheridae (pea crabs) are almost always small (less than 1 cm), pale parasites or commensals found on or in bivalve mollusks (especially oysters), tube-dwelling polychaetes, and other invertebrates. The square-backed crabs (Grapsidae) [Fig. 13] are characteristic of warm, marshy areas and rocky shores. They mark the trend toward the true land crabs (Gecarcinidae) [Fig. 15*c*], the depredations and intrusions of which are familiar to all who live in the tropics. Similar to gecarcinids, members of the family Ocypodidae (Fig. 15*b*) occupy burrows during the day or high tide. Ghost crabs (*Ocypode*) scuttle almost unseen over sandy beaches and dunes at night. Fiddler crabs (*Uca*) emerge at low tide, often in large numbers, to feed on mud and sand deposited in embayments; males also engage in courtship and combat displays using a greatly inflated major chela. *See* ARTHROPODA; CRUSTACEA.

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Decca

A hyperbolic radio navigation system. The Decca Navigator hyperbolic radio-navigation system was invented in the United States in 1937 and developed in 1944 by the British Decca Company. More than 50,000 ships of all classes, including a large and growing number of recreational craft, use the system, and about 750 helicopters and fixed-wing aircraft; coverage is shown in **Fig. 1**.

The system employs radio transmitting stations grouped in chains of four (in a few cases, three) established at precisely known positions. A chain comprises a central master station with outlying slave stations, all radiating unmodulated continuous-wave signals on spot frequencies between 70 and 130 kHz. In a given chain, each of the slave transmissions is locked in a known phase relationship with the master. Assuming a constant speed of signal propagation, the phase difference between the signals received from the master and a given slave is a measure of the difference in the distances from the receiver to the two stations. The phase difference can thus locate the receiver on a line of position in the form of a hyperbola with the stations as foci. The observed phase difference between the master and the signal from a second slave similarly identifies an intersecting line of position, fixing the position.

In its simplest form, the on-board receiver displays the phase differences as numerical line-of-position readings, which the user plots on a marine chart overprinted with a lattice of correspondingly numbered hyperbolic curves. A four-station chain produces three overlapping line-of-position patterns, printed on charts in red, green, and purple. For manual plotting (**Fig. 2**), the charts normally carry the two patterns cutting at the best angle and omit the third. Computer-based Decca receivers can make use of all three patterns to form the fix. Such receivers also convert the hyperbolic fix into latitude and longitude, position with respect to a waypoint, and other desired information.

Lanes and lane identification. The line-of-position readings are expressed in units and fractions of a lane, where a lane is the space corresponding to a change of 360° in the master–slave phase difference. Typically the receiver measures the phase difference to 3.6° (0.01 lane), corresponding to a line-of-position resolution of about 16 ft (5 m) on the

master-to-slave baselines where the lane width is minimum. The measurement is, however, ambiguous, since there is no way of determining directly the number of whole cycles in the phase difference. Accordingly the receiver automatically keeps count of the successive lanes that it passes through, once the correct lane numbers at the starting position have been inserted. These numbers are supplied periodically by additional transmissions which identify, for each pattern, the correct lane within a relatively large group of lanes known as a zone. The zones are about 6 mi (10 km) wide on the baselines, so that in practice there is a spacing of 6 mi (10 km) or more between ambiguous lines of position. The lane identification values are either displayed as numerical readings or, in computer-based receivers, processed automatically.

Frequencies. Each chain is assigned a set of four frequencies (see **table**), which are multiples of a fundamental value f between 14.0 and 14.4 kHz. The harmonic relationship allows the receiver to derive, by frequency multiplication, a common frequency from the master and a given slave so that their phase can be compared. For example, master $6f \times 3$ and green slave $9f \times 2$ are each equal to $18f$, and the green line-of-position pattern is the result of phase comparison at that frequency.

For lane identification, each station transmits in turn, in a cycle repeated every 20 s, a half-second burst comprising the four phase-coherent harmonics $5f$, $6f$, $8f$, and $9f$. The other stations cease transmitting on these frequencies for the half-second. From the fourfold signals, the receiver derives, by one of several methods depending on the software, a high-integrity signal having the fundamental frequency f. Comparing master and slave phases at that frequency generates what amounts to a coarse

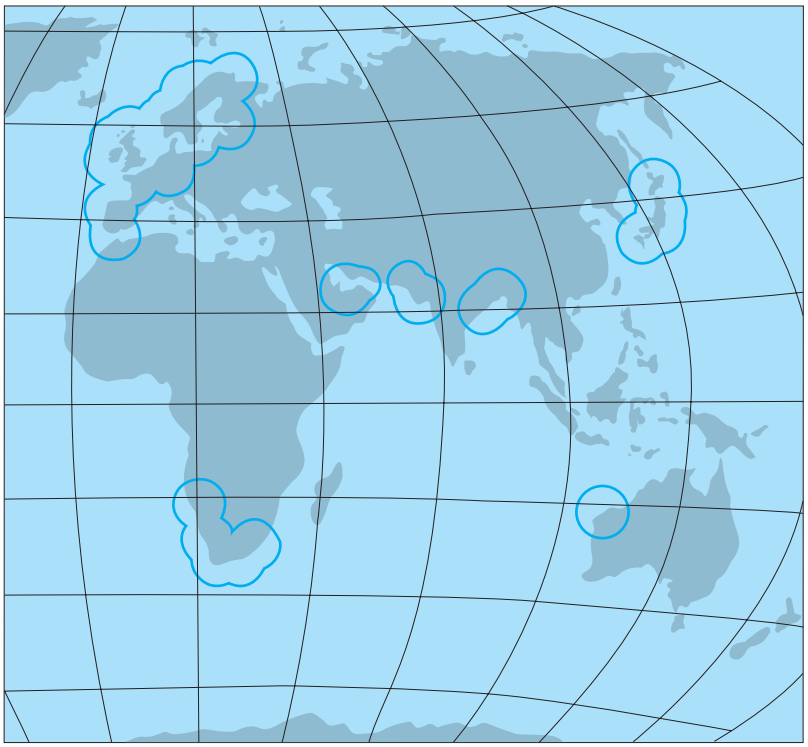

Fig. 1. Decca coverage areas. (*After Racal-Decca Marine Navigation Ltd, Decca Navigator Coverage Diagrams, 1988*)

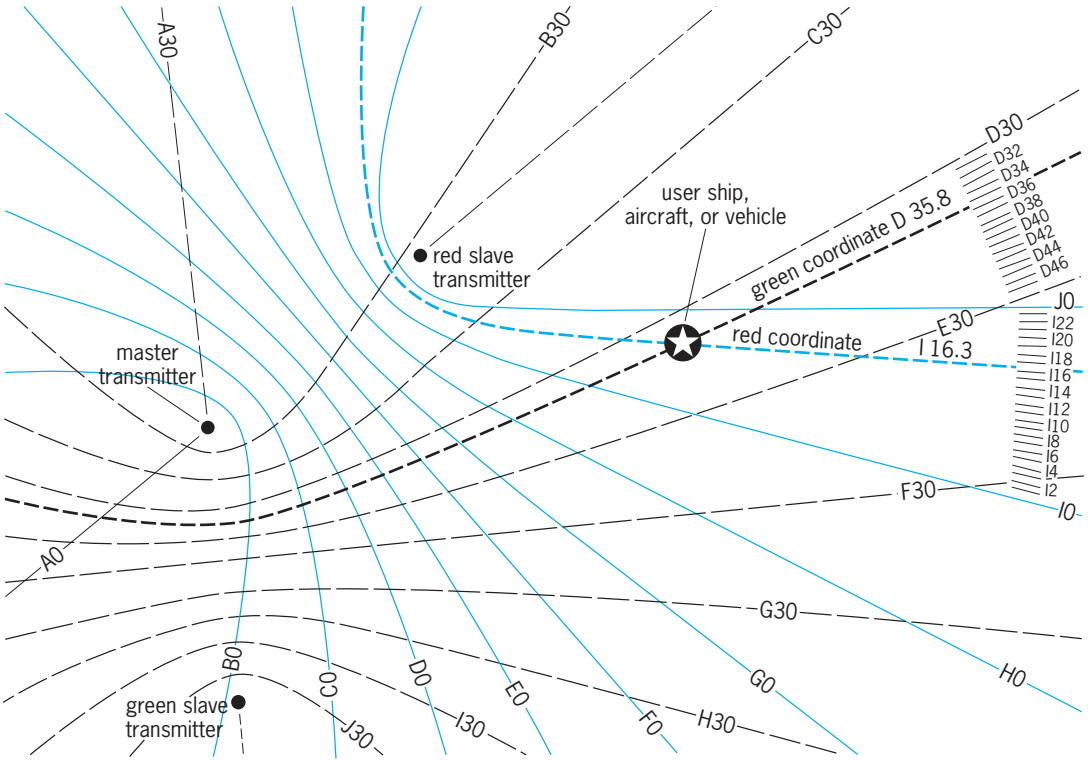

Fig. 2. Fix plotted from line-of-position readings Red I 16.30, Green D 35.80, where I and D denote the zones; lanes are shown at top right. (*After C. Powell, Radio Position Fixing for Yachtsmen, Adlard Coles Ltd, London, 1986*)

Decca transmitted and comparison frequencies*

	Frequency, kHz	Harmonic, *f*	Lane or zone width on baseline, ft (m)
Transmitted frequencies			
Master station	85.0000	6	
Purple slave	70.8333	5	
Red slave	113.3333	8	
Green slave	127.5000	9	
Phase-comparison frequencies			
Purple lanes pattern	425.00	30	1,155 (352.1)
Red lanes	340.00	24	1,443 (440.1)
Green lanes	255.00	18	1,925 (586.8)
Zones (all patterns)	14.1666	1	34,652 (10,562.0)

*For Decca chain 05B.

hyperbolic pattern superimposed upon the lanes and serving to identify the correct lane for each pattern, assuming that the zone is known. Some computer-based receivers treat the continuous and lane identification signals in a relationship that differs from the foregoing basic description.

Range and accuracy. The range of a Decca chain is typically some 240 nautical miles (440 km) by night and about twice that distance by day. Repeatability of position fixing is as high as a few tens of meters in favorable coverage during summer daytime, falling to a few miles at night near the limit of range. The variation with time occurs through contamination of the wanted ground-wave component of the received signals by the sky wave reflected by the ionosphere, which is strongest at night and varies randomly in phase. Published accuracy data include predicted levels of repeatability (random position-fixing error), taking seasonal and diurnal sky-wave propagation variations into account. The composite lane identification signal is relatively insensitive to the effects of sky-wave contamination, as a result of the fourfold frequency diversity of the transmission. Data are also published on fixed (systematic) position-fixing errors observed in coastal areas. *See* ELECTRONIC NAVIGATION SYSTEMS; HYPERBOLIC NAVIGATION SYSTEM; MARINE NAVIGATION. Claud Powell

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Decibel

A logarithmic unit used to express the magnitude of a change in level of power, voltage, current, or sound intensity. A decibel (dB) is $^1/_{10}$ bel.

In acoustics a step of 1 bel is too large for most uses. It is therefore the practice to express sound intensity in decibels. The level of a sound of intensity I in decibels relative to a reference intensity I_R is given by notation (1). Because sound intensity is proportional to the square of sound pressure P, the level in decibels is given by Eq. (2). The reference pressure is usually taken as 0.0002 dyne/cm^2 or 0.0002 microbar. (The pressure of the Earth's atmosphere at sea level is approximately 1 bar.) A sinusoidal alternation in pressure at a frequency of 1000 Hz is barely audible to the average person when it has a root-mean-square sound pressure of 0.0002 microbar. By this definition such a tone has a sound pressure level of 0 dB. *See* SOUND PRESSURE.

$$10 \log_{10} \frac{I}{I_R} \qquad (1)$$

$$10 \log_{10} \frac{P^2}{P_R{}^2} = 20 \log_{10} \frac{P}{P_R} \qquad (2)$$

The neper is similar to the decibel but is based upon natural (napierian) logarithms. One neper is equal to 8.686 dB. *See* NEPER; VOLUME UNIT (VU).

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Deciduous plants

Woody higher plants that regularly lose their leaves at the end of each growing season. Dropping of the leaves occurs at the inception of an unfavorable season characterized by either cold or drought or both. Many woody plants of temperate climates have the deciduous habit, and it may also occur in those of tropical regions having alternating wet and dry seasons. Many deciduous trees and shrubs of regions with cold winters become evergreen when grown in a warm climate. Conversely, such trees as magnolias, evergreen in warm areas, become deciduous when grown in colder climates. *See* LEAF; PLANT PHYSIOLOGY; PLANT TAXONOMY. Nelle Ammons

Decision analysis

An applied branch of decision theory. Decision analysis offers individuals and organizations a methodology for making decisions; it also offers techniques for modeling decision problems mathematically and finding optimal decisions numerically. Decision models have the capacity for accepting and

quantifying human subjective inputs: judgments of experts and preferences of decision-makers. Implementation of models can take the form of simple paper-and-pencil procedures or sophisticated computer programs known as decision aids or decision systems.

Principles and methodology. The methodology is rooted in postulates of rationality—a set of properties which preferences of rational individuals must satisfy. One such property is transitivity: if an individual prefers action a to action b and action b to action c, he or she should prefer a to c. From the rationality postulates, principles of decision-making are derived mathematically. The principles prescribe how decisions ought to be made, if one wishes to be rational. In that sense, decision analysis is normative.

The methodology is broad and must always be adapted to the problem at hand. An illustrative adaptation to a class of problems known as decision-making under uncertainty (or risk) and consists of seven steps (**Fig. 1**):

1. The problem is structured by identifying feasible actions $a_1, \ldots, a_k, \ldots, a_n$, one of which must be decided upon; possible events $\omega_1, \ldots, \omega_j, \ldots, \omega_m$, one of which occurs thereafter; and outcomes $z_{11}, \ldots, z_{kj}, \ldots, z_{nm}$, each of which results from a combination of decision and event. Problem structuring can be facilitated by displays such as decision trees, decision matrices, and influence diagrams.

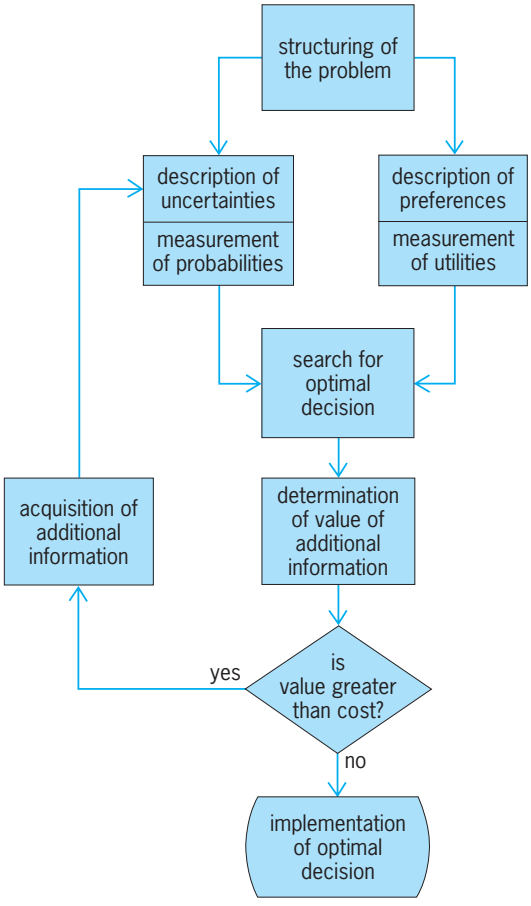

Fig. 1. A methodology of decision analysis.

2. At the time of decision-making, the event that will actually occur cannot be predicted perfectly. The degree of certainty about the occurrence of an event, ω_j, given all information at hand, is quantified in terms of the probability of the event, $P(\omega_j)$. *See* PROBABILITY.

3. Preferences are personal: the same outcome may elicit different degrees of desirability from different individuals. The subjective value that a decision-maker attaches to an outcome is quantified and termed the utility of outcome, $u(z_{kj})$.

4. The preceding steps conform to the principle of decomposition: probabilities of events and utilities of outcomes must be measured independently of one another. They are next combined in a criterion for evaluating decisions. The utility of a decision $U(a_k)$ is defined in Eq. (1) as the expected utility of

$$U(a_k) = \sum_{j=1}^{m} u(z_{kj}) P(\omega_j) \tag{1}$$

the outcome. The optimal, or the most preferred, decision a^* is one with the maximum utility, as given in Eq. (2), where the symbol max indicates the operation of finding the maximum by varying the index k from 1 to n. Relations (1) and (2) constitute the principle of utility maximization.

$$U(a^*) = \max_{k} U(a_k) \tag{2}$$

5. The probability $P(\omega_j)$ encodes the current state of information about a possible event. Often, additional information can be acquired in the hope of reducing the uncertainty. The monetary value of such information is computed before purchase and compared with the cost of information. Thus, one can determine whether or not acquiring information is economically rational.

6. The source of information may be a real-world experiment, a laboratory test, a mathematical model, or the knowledge of an expert. The informativeness of the source is described in terms of a probabilistic relation $P(I|\omega_j)$ between information I and event ω_j. This relation, known as the likelihood function, makes it possible to revise the prior probability $P(\omega_j)$ of the event and to obtain a posterior probability $P(\omega_j|I)$ of the event, conditional on additional information. The revision is carried out via Bayes' rule, as shown in Eq. (3).This rule (**Fig. 2**) is also known

$$P(\omega_j|I) = \frac{P(I|\omega_j)\,P(\omega_j)}{\sum_{j=1}^{m} P(I|\omega_j)\,P(\omega_j)} \tag{3}$$

as the Bayesian learning model. *See* BAYESIAN STATISTICS; ESTIMATION THEORY.

7. Given the additional information, prior probabilities can be replaced by posterior probabilities, and the analysis can be repeated from step 4 onward. Steps 4–6 may be cycled many times, until the cost of additional information exceeds its value, at which moment the optimal decision is implemented.

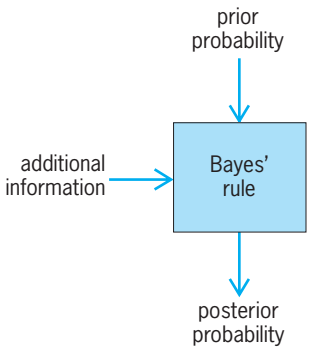

Fig. 2. Model for revising the state of information or knowledge.

Situations and techniques. A spectrum of technique exists for modeling and solving various tasks within a decision analysis; these tasks vary with situations. Six elements of decision situations are highlighted below along with the associated modeling tasks.

1. The decision-maker may be either an individual or a group, in which case utility functions of group members must be aggregated.

2. The stakeholder (one who is directly affected by the outcome of a decision) may be the decision-maker, another individual, or a group; a decision-maker may wish to include in the analysis a utility function of the stakeholder in addition to his or her own.

3. Events leading to outcomes either may be known with certainty or be uncertain at the time of decision-making. Accordingly, decision-making under conditions of certainty and uncertainty (or risk) requires different techniques.

4. Information about possible events may be acquired from single or multiple sources; in the latter case, information must be aggregated.

5. Outcomes may be one-dimensional (for example, monetary damage caused by a flood) or multidimensional (for example monetary damage and lives lost); they may be quantitative or qualitative (for example, verbal descriptors of product quality or the state of one's health). Outcome dimensions are often called attributes, and a decision-maker's preferences are described in terms of a single-attribute or multiattribute utility function. The latter is usually a composition of single-attribute utility functions and parameters that encode preferred trade-offs between outcome dimensions.

6. Time scale: decision problems are static when one decision must be made at a specified time; they are dynamic when the optimal timing of action must be decided (a stopping problem) or when many decisions must be made at predetermined times (a sequential decision problem). Dynamic problems may have a finite or infinite horizon, and they require two additional models. First, it is necessary to describe a sequence of events as a stochastic process. Second, it is necessary to describe preferences for sequences of outcomes; this may involve discounting the value of future outcomes relative to the present ones. *See* STOCHASTIC PROCESS.

Multiobjective decision analysis. A static decision problem under conditions of certainty and with multiple objectives is structured as follows. Objectives of import to a decision-maker are identified (for example, maximize economic efficiency, maximize customer satisfaction, minimize environmental impact). To measure the attainment of the objectives, attributes $z_1, \ldots, z_i, \ldots, z_r$ are defined, one or more for each objective. For each feasible action a_k, outcomes $z_{k1}, \ldots, z_{ki}, \ldots, z_{kr}$, which represent levels of the attributes, are determined.

Suppose that for each attribute a higher outcome is preferred to a lower outcome. In a comparison of two actions, action a_k is said to dominate action a_l if $z_{ki} \geq z_{li}$ for $i = 1, \ldots r$, with at least one strict inequality. Any action dominated by some other action is eliminated from further analysis. To rank the remaining, nondominated actions, preferences of the decision maker must be modeled.

The subjective values that the decision maker attaches to multidimensional outcomes are quantified in terms of a multiattribute utility function. The simplest form is additive, as shown in Eq. (4). Here,

$$U(a_k) = \sum_{i=1}^{r} \lambda_i u_i(z_{ki}) \tag{4}$$

the utility of an action $U(a_k)$ equals the sum of scaled single-attribute utilities $u_i(z_{ki})$, where the scaling constants λ_i (also termed the trade-off coefficients) quantify the trade-offs preferred by the decision maker in the following sense: λ_i is the utility of an action that yields the most desirable outcome in attribute z_i and the least desirable outcomes in the remaining attributes. By ordering the utilities of actions $U(a_k)$, $k = 1, \ldots, n$, from the highest to the lowest, the most preferred ranking of actions is decided.

Measurement of probability and utility. Measurement of probability and utility functions is guided by principles of decision theory, statistical estimation procedures, and empirical laws provided by behavioral decision theory—a branch of cognitive psychology.

Statistical techniques for estimating probabilities are applicable when the available information consists of past observations of events, outputs from models simulating behavior of a system that generates the events, or other numerical data. When information is in the form of qualitative statements and images or general knowledge of experts, probabilities are assessed via human judgment. For instance, meteorologists assess probabilities of precipitation occurrence, and physicians assess probabilities of correctness of their diagnoses. Several questions arise: How do people perform such judgmental tasks? How can they be aided in these tasks? How can their performance be evaluated? How can they be trained to become good probability assessors?

These questions are studied within behavioral decision theory. Although far from complete, the

theory has gathered vast empirical knowledge about human judgmental capabilities. When processing information intuitively, people rely on heuristics—simple subjective rules for inference. Heuristics are at best approximations to normative inference rules and frequently lead to predictable biases. For instance, a heuristic known as anchoring and adjustment operates as follows: Information first received is used to form a prediction, which is next adjusted in light of new information. Because adjustments are usually insufficient, the performance of an expert does not increase commensurately with information. Yet, the expert's confidence in his prediction increases steadily. At the end of information processing, most experts are overconfident.

Understanding heuristics and biases is important for designing probability measurement protocols that lessen erroneous predictions. Methods exist for evaluating the performance of probabilistic predictions and providing experts with feedback. Empirical evidence suggests that possessing professional expertise does not assure good probability assessment. Training in probabilistic thinking usually improves an expert's performance.

Techniques for measuring utilities are numerous. Most often, a decision-maker is presented with two simple alternatives, and his task is to choose. For example, the choice may be between an alternative that yields outcome y with certainty and an alternative that yields outcome x with probability p or outcome z with probability $1 - p$, a gamble. Preferences expressed in such simple situations supply information for constructing a utility function. A decision model (1)–(2) with this utility function applied to simple situations should replicate the decision-maker's choices. In that sense, the decision model should be descriptive.

The main premise of the analysis, however, is that the model will be applied to complex situations, and will prescribe decisions that are optimal according to the preferences encoded in the utility function. Thus, the model guarantees that all decisions, in simple and complex situations, satisfy the postulates of rationality, that is, all decisions are coherent.

Experimental evidence points to the limited human capacities for making coherent decisions intuitively, particularly in situations with uncertainties and multidimensional outcomes. Human preferences can easily be affected by the format of information display, the ways questions are framed and sequenced, and other factors which, from the normative point of view, should not matter. Research is in progress toward a better understanding of cognitive processes underlying preference formation and intuitive decisions. The ultimate objective is to improve techniques for interfacing a decision analysis with a human decision-maker whom the analysis seeks to serve. *See* COGNITION; DECISION THEORY; OPERATIONS RESEARCH; OPTIMIZATION; SYSTEMS ANALYSIS; SYSTEMS ENGINEERING.

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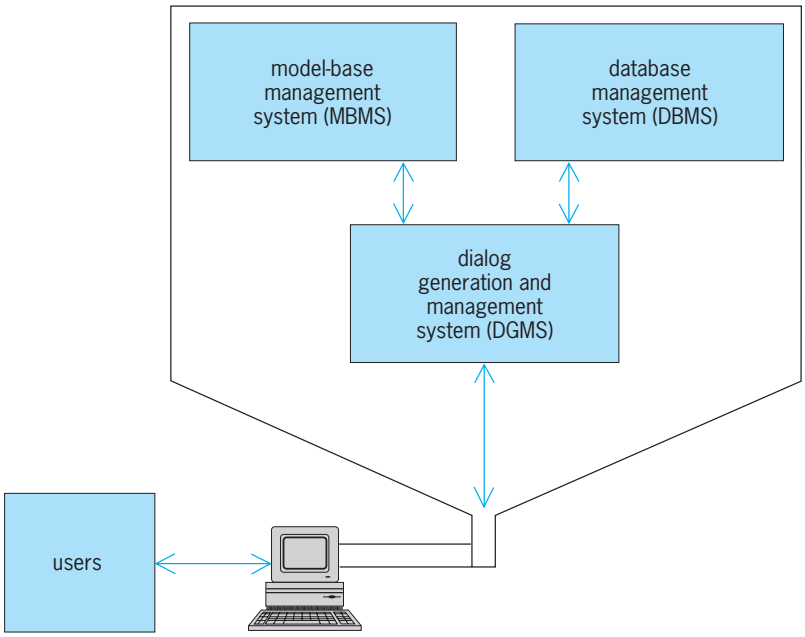
Decision support system

A system that supports technological and managerial decision making by assisting in the organization of knowledge about ill-structured, semistructured, or unstructured issues. A structured issue has a framework comprising elements and relations between them that are known and understood. Structured issues are generally ones about which an individual has considerable experiential familiarity. A decision support system (DSS) is not intended to provide support to humans about structured issues since little cognitively based decision support is generally needed for these experientially familiar situations. Artificial-intelligence based expert systems are very appropriate for these tasks.

Scope of application. The primary components of a decision support system are a database management system (DBMS), a model-base management system (MBMS), and a dialog generation and management system (DGMS) [see **illus.**]. Emphasis in the use of a decision support system is upon provision of support to decision makers in terms of increasing the effectiveness of the decision-making effort. This support involves the systems engineering steps of formulation of alternatives, the analysis of their impacts, and interpretation and selection of appropriate options for implementations. Efficiency in terms of time required to evolve the decision, while important, is usually secondary to decision effectiveness.

Types of decisions and systems. Decisions may be described as structured or unstructured, depending upon whether or not the decision-making process can be explicitly described prior to its execution. Generally, operational performance decisions are more likely than strategic planning decisions to be prestructured. Thus, expert systems are usually more appropriate for operational performance and operational control decisions, while decision support systems are more appropriate for strategic planning and management control. Most expert systems are based upon extensive use of what are generally called production rules. This rule-based reasoning may not cope well with situations that require either formal knowledge-based reasoning or skill-based expert reasoning. *See* EXPERT SYSTEMS.

Systems engineering steps. Fundamental to the notion of a decision support system is the formal assistance provided to humans in assessing the situation of potential interest, identifying alternative courses of action, formulating the decision situation, structuring and analyzing the decision situation in terms of obtaining the impacts of alternative potential



Components of a decision support system (DSS).

courses of action, and interpreting the results of analysis of the alternatives in terms of the value system of the decision maker. These are the basic formulation, analysis, and interpretation steps of systems engineering in expanded form. *See* SYSTEMS ENGINEERING.

Database management system. An appropriate database management system must be able to work with both data that are internal to the organization and data that are external to it. If there are multiple decision makers, then personal databases, local databases, and systemwide databases are generally needed. A database management system should be able to cope with a variety of data structures that allow for probabilistic, incomplete, and imprecise data, and data that are unofficial and personal as contrasted with official and organizational. It should also be capable of informing the support-system user of the types of data that are available and how to gain access to them. *See* DATABASE MANAGEMENT SYSTEM.

Model-base management system. The need to provide recommendation capability in a decision support system leads to the construction of model-base management systems. Such systems provide sophisticated analysis and interpretation capability in a decision support system. A model-base management system enables the decision maker to explore the decision situation through use of the database by a model base of algorithmic procedures and associated model management protocols. This process may involve use of modeling statements in some procedural or nonprocedural language; model subroutines, such as mathematical programming packages, that are called by a management function; and data abstraction models. This last approach is close to the expert-system approach in that there exist element, equation, and solution procedures that together make up an inference engine. Advantages include ease of updating and use of the model for explanatory and explication purposes.

The model-base management system should provide flexibility upon system-user request through a variety of prewritten models, such as linear programming and multiattribute decision analysis models, and procedures to use these models. The system should also allow for the development of user-built models and heuristics that are developed from established models. It should also be possible to perform sensitivity tests of model outputs, and to run models with a range of data in order to obtain the response to a variety of what-if questions. *See* DECISION ANALYSIS; LINEAR PROGRAMMING; MODEL THEORY.

Dialog generation and management. The dialog generation and management system is designed to satisfy knowledge representation, and control and interface requirements of the decision support system. It is responsible for presentation of the information outputs of the database management system and model-base management system to the decision makers, and for acquiring and transmitting their inputs to these two systems. Specifically, it is responsible for producing the output representations of the decision support system, for obtaining the decision-maker inputs that result in the operations on the representations, for interfacing to the memory aids, and for explicit provision of the control mechanisms that enable the dialog between user input and output and the database management system and model-base management system. *See* DECISION THEORY.

Andrew P. Sage

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Decision theory

A broad range of concepts which have been developed to both describe and prescribe the process of decision making, where a choice is made from a finite set of possible alternatives. Normative decision theory describes how decisions should be made in order to accommodate a set of axioms believed to be desirable; descriptive decision theory deals with how people actually make decisions; and prescriptive decision theory formulates how decisions should be made in realistic settings. Thus, this field of study involves people from various disciplines: behavioral and social scientists and psychologists who generally attempt to discover elaborate descriptive models of the decision process of real humans in real settings; mathematicians and economists who are concerned with the axiomatic or normative theory of decisions; and engineers and managers who may be concerned with sophisticated prescriptive decision-making procedures.

Concepts and terminology. Decisions have various degrees of importance in the life of an individual, a group, or an organization. The importance of a decision depends upon its consequences or outcomes as determined by: (1) controlled factors, represented by the alternative courses of action or alternate strategies available to the decision maker; and (2) uncontrolled factors, or states of nature, which interact with the strategies.

A decision is an allocation of resources. It is generally irrevocable, although a later decision may sometimes reverse the effects of an initial decision. The decision maker is the person who has authority over the resources that are being allocated. This person makes a decision in order to further the achievement of some objective which is felt to be desirable.

A decision maker should make decisions that are consistent with values, or those things that are important to the decision maker, or those parties represented by the decision maker. The most common value is economic, the effort to increase wealth; but such personal or social values as happiness, security, and equity are also important. Sometimes a distinction is made between a goal and an objective. A goal refers to a specific degree of satisfaction of a given objective. For example, the objective of a decision to exercise or diet might be to lose weight, and the goal might be to weigh 10 lb (4.5 kg) less in a month.

The important features of a decision situation are (1) a decision to be made and course of action to be taken, (2) the unknown events and outcomes which can affect the result, and (3) the obtained result. Decision theory is based on construction of models which represent logical mathematical representations of the relationships among these three features. Use of these models leads to estimation of the possible implications of each alternative course of action that might be taken, and thereby to understanding of the relationships between actions and objectives, and to selection of the most appropriate decision.

At the time a decision is made or taken, the decision maker has at least two alternatives or courses of action (otherwise, no decision is possible). When an alternative is chosen and committed to, the decision has been made or taken, and uncertainties (uncontrollable chance elements) generally arise. Different alternatives may be associated with different uncertainties. It is the combination of alternatives with uncertainties that ultimately leads to an outcome. Thus, an outcome is a generally uncertain result that follows from the decision situation, whose impacts should be measured on the decision maker's value scale.

A simple decision is a situation in which there is only one decision to be made, even though there might be many alternatives. If there are two or more decisions to be made, a problem of strategy arises. Sometimes, several decisions need to be made at the same time, and this special case is called the portfolio problem. Often, the various decisions associated with the strategy are of a similar nature, and there are not sufficient resources to enable selection of all alternatives. The prototypical example is an investment portfolio. Here, the decision maker is aware of a number of investments that might potentially be made, but there is a finite budget. In such a situation, the decision maker must accomplish a prioritization of the various opportunities. If one opportunity has a higher priority than another, the decision maker would prefer to invest in the alternative that leads to this opportunity. In other cases, sequential decisions are involved. Here, the decision maker is able to observe the outcome that results from one decision before having to make a subsequent decision.

The consequences of the chosen strategies are dependent quantities commonly termed outcomes. These are often measured by a variety of quantitative or qualitative parameters, such as the number of dollars returned from an investment, or the quality of life that results from a successful medical procedure, and by the probability of the various possible consequences following a selected course of action. The aggregate outcome that follows from a given decision is known as the expected relative value or expected utility.

Categories of problems. Efforts in decision theory may be divided into five categories:

1. Decision under certainty issues are those in which each alternative action results in one and only one outcome and where that outcome is sure to occur. The decision situation structural model is established and is correct. There are no parametric uncertainties.

2. Decision under probabilistic uncertainty issues are those in which one of several outcomes can result from a given action depending on the state of nature, and these states occur with known probabilities. The decision situation structural model is established and is correct. There are outcome uncertainties, and the probabilities associated with these are known precisely. The utility of the various event outcomes can be precisely established. *See* PROBABILITY.

3. Decision under probabilistic imprecision issues are those in which one of several outcomes can result from a given action depending on the state of nature, and these states occur with unknown or imprecisely specified probabilities. The decision situation structural model is established and is correct. There are outcome uncertainties, and the probabilities associated with the uncertainty parameters are not all known precisely. The utility of the event outcomes can be precisely established.

4. Decision under information imperfection issues are those in which one of several outcomes can result from a given action depending on the state of nature, and these states occur with imperfectly specified probabilities. The decision situation model is established but may not be fully specified. There are outcome uncertainties, and the probabilities associated with these are not all known precisely. Imperfections in knowledge of the utility of the various event outcomes may exist as well.

5. Decision under conflict and cooperation issues are those in which there is more than a single decision maker, and where the objectives and activities of

one decision maker are not necessarily known to all decision makers. Also, the objectives of the decision makers may differ.

Problems in any of these groupings may be approached from a normative, descriptive, or prescriptive perspective. Problems in category 1 are those for which deterministic principles may be applied. This condition of known states of nature ignores the overwhelming majority of issues faced in typical realistic decision issues.

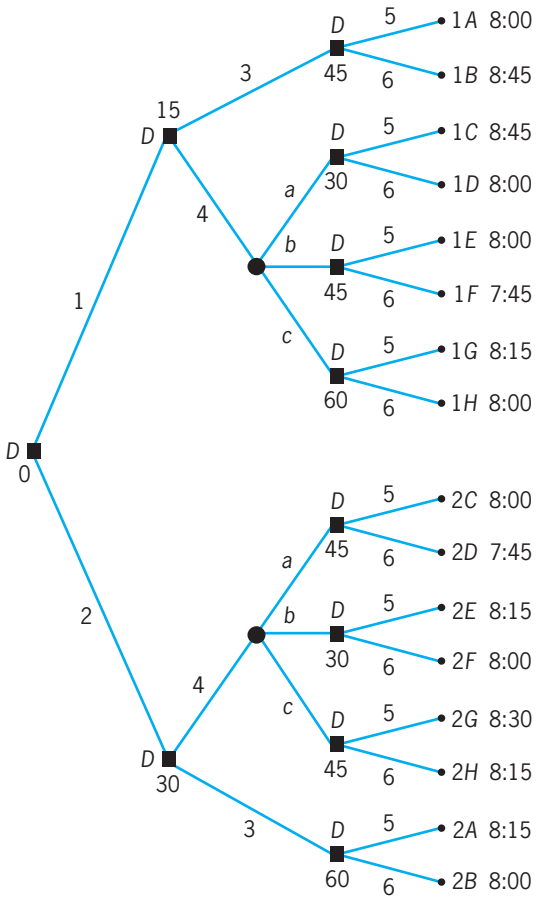
Problems in category 5 are game theory–based problems, and will not be considered here. Most decision theory developments are concerned with issues in category 2. *See* GAME THEORY.

The basic normative decision problem in category 2 is conceptually straightforward: From a set of alternative strategies (S_n), choose the one which when combined with the relevant states of nature (N_j) will produce the outcomes (O_{nm}) which best satisfy the goals or objectives of the decision maker. Some type of penalty or benefit criterion, or multiple criteria, are used to evaluate the utility of the outcome O_{nm}. Normative decision theory implies that all four factors should be known to make a completely rational and meaningful decision. Behavioral or descriptive decision theory tries to explain why real people often makes seemingly irrational decisions.

Sample problem. By considering the relatively simple decisions which typically have to be made before getting to work, the possible difficulties can be demonstrated. In **Fig. 1**, it is assumed that the decision maker has three basic decisions to make on the way to work: (1) to get up at 7:00 A.M. (choice 1) or 7:15 A.M. (choice 2), with the associated consequence that after dressing he or she would depart at 7:15 or 7:30 A.M., respectively; (2) to choose routes A or B (choices 3 and 4) with the expected outcomes that route A always takes 30 min while route B, resulting from choice 4, can take 15, 30, or 45 min (*a*, *b*, or *c*), depending on traffic conditions; and (3) to stop for breakfast, which takes 15 min, or skip breakfast (choices 5 and 6).

The decision sequence is illustrated (Fig. 1) in the form of a decision tree. It shows all the possible intermediate and final outcomes which would result from all possible choices of alternatives posed by this example. The decision nodes at which decisions have to be made are marked *D*.

Several interesting observations can be made about the decision tree. (1) Although there are 16 possible outcomes, there are only 8 possible initial strategies since the decision outcomes *a*, *b*, and *c* are mutually exclusive and determined by the state of nature, traffic congestion, at the time. (2) The decision problem has been bounded by enumerating all admissible strategies, states of nature, and outcomes. This has already involved some judgment in excluding other strategies, states of nature, and outcomes on the basis that they are unlikely to occur, or because the decision maker is specifically interested only in the alternatives presented. (3) Although the alternative strategies and outcomes are readily apparent from this particular decision tree, the



Key to decisions:

1 = Get up and dress when alarm rings
2 = Sleep 15 min more, then get up and dress
3 = Choose route A to work
4 = Choose route B to work
5 = Eat breakfast
6 = Skip breakfast

Fig. 1. Decision tree for the problem given in the text. Numbers at nodes represent time past alarm (7:00 A.M.), and numbers at branches refer to the decision mode or state of nature.

fourth ingredient of a decision problem, the objectives of the decision maker, have not yet been defined. The objectives which generate the criteria for evaluating the decision problem represent the critical link in reaching the best decision for that particular decision maker at this particular time. In Fig. 1 it is clear that the decision relative to choice of best strategy will depend on the decision maker's objectives. These objectives are translated into values, to a related concept called utility, and to the expected value of the utilities in most decision theory–based approaches.

If the objective is to definitely arrive at work at 7:45 A.M. sharp, it is possible to immediately eliminate all strategies which do not have 7:45 A.M. as their final outcome as inadmissible strategies. In this case the only admissible strategy would be (1-3-6) since any other strategy often would result in coming to work late.

If, however, the decision maker decides that breakfast is a vital function and will always insist on breakfast, all strategies which have a branch 6 in the decision tree will be eliminated. If the objective is still to get to work as early as possible, the decision must now be made between strategy (1-3-5) and strategy (1-4-5), which may under certain circumstances result in arrival at work earlier than (1-3-5), that is, outcome 1*C*. These two strategies dominate all other admissible strategies because under all circumstances they will result in arrival at work earlier than any of the other strategies. Thus, when one strategy shows dominance over some other strategies, all but the dominant strategies can be eliminated from consideration. Strategies (1-3-5) and (1-4-5) do not dominate each other since under certain circumstances one or the other will have the more desirable outcome.

The choice of strategy (1-3-5) leads to a decision-under-certainty outcome since the outcome for this strategy is uniquely predictable and all states of nature are known. The outcome of strategy (1-4-5) depends on the traffic conditions (*a*, *b*, or *c*) or states of nature which are not under the decision maker's control. If no information is available about the likelihood or probability of occurrence of states *a*, *b*, or *c*, the decision maker will be faced with decision making under probabilistic imprecision. If the probability of the states of nature *a*, *b*, or *c* can be predicted from past experience or from observations, a probability can be associated with each of the possible outcomes. This represents decision making under probabilistic uncertainty if it is assumed that the values of the various outcomes are also known.

Value or utility theory. Except in very well defined decision problems in which the outcomes are clearly represented by such an objective parameter as dollars or time, there are conflicting and incommensurable objectives which require subjective value judgments in order to determine the value or utility of each objective. This inescapable aspect of decision theory has led to the development of value or utility theory. Utility theory strives to evaluate the subjective relative value of outcomes to individuals and groups. For example, the utility or value of $100,000 to a multimillionaire in a gambling decision is obviously quite different from its value to a person whose total possessions amount to $100,000. Similarly, the question of the relative values of 15 min extra sleep versus no breakfast versus coming to work late requires some resolution before the objectives and evaluation criteria for the prior decision problem can be established.

Bases of normative decision theory. The general concepts of axiomatic or normative decision theory formalize and rationalize the decision-making process. The major contributions of decision theory are to structure the decision problem and to define the scope and importance of components so that each part of the decision can be evaluated before synthesizing the available information into a rational decision.

Normative decision theory depends on the following assumptions:

1. Past preferences are valid indicators of present and future preferences.
2. People correctly perceive the values of the uncertainties that are associated with the outcomes of decision alternatives.
3. People are able to assess decision situations correctly, and the resulting decision situation structural model is well formed and complete.
4. People make decisions that accurately reflect their true preferences over the alternative courses of action, each of which may have uncertain outcomes.
5. People are able to process decision information correctly.
6. Real decision situations provide people with decision alternatives that allow them to express their true preferences.
7. People accept the axioms that are assumed to develop the various normative theories.
8. People make decisions without being so overwhelmed by the complexity of actual decision situations that they would necessarily use suboptimal decision strategies.

Given these necessary assumptions, there will exist departures between normative and descriptive decision theories. A principal task of those aiding others in decision making is to retain those features from the descriptive approach which enable an acceptable transition from normative approaches to prescriptive approaches. The prescriptive features should eliminate potentially undesirable features of descriptive approaches, such as flawed judgment heuristics and information processing biases, while retaining acceptable features of the normative approaches. *See* DECISION ANALYSIS; DECISION SUPPORT SYSTEM.

Decision matrices. The information necessary for a normative decision can be summarized in a decision matrix (**Fig. 2**). The alternative strategies, pure (independent from each other) or mixed (combinations of pure strategies), are listed in the first column. These strategies (S_1 to S_n) can be generated by the decision maker, or by an advisor to the decision maker often called a decision analyst, and should contain all possible courses of action which the decision maker can identify as feasible.

Across the top row, all possible states of nature (N_1 to N_m) which might affect the outcome of the various strategies are enumerated. The decision matrix can become quite large if, for example, it is to represent a complete decision tree, such as the tree of Fig. 1. In that case, the trade-off between time and money expended on a decision and the value of a more or less reliable decision can become a decision problem in itself. There are computational approaches for resolution of decision trees that involve procedures from dynamic programming and optimization theory, and often involve such computational strategies as "average out and fold back." *See* OPTIMIZATION.

The probabilities of occurrence of various states of nature (p_1 to p_m) must be estimated and listed with the states of nature. These critical values should be as reliable as possible. In value theory a distinction is made between the subjective probability of

States of nature		N_1	N_2	— —	N_j	— — — — — —	N_m	Expected value
Probabilities of occurrence		p_1	p_2	— —	p_j	— — — — — —	p_m	
Alternate strategies	S_1	O_{11}	O_{12}	— —	O_{1j}	— — — — — —	O_{1m}	EV_1
	S_2	O_{21}	O_{22}	— —	O_{2j}	— — — — — —	O_{2m}	EV_2
	|	|	|		|		|	|
	|	|	|		|		|	|
	S_i	O_{i1}	O_{i2}	— —	O_{ij}	— — — — — —	O_{im}	EV_i
	|	|	|		|		|	|
	|	|	|		|		|	|
	|	|	|		|		|	|
	S_n	O_{n1}	O_{n2}	— —	O_{nj}	— — — — — —	O_{nm}	EV_n

Fig. 2. Form of a decision matrix. O_{ij} = outcome corresponding to the interaction of a strategy (S_i) and a state of nature (N_j).

occurrence which reflects the decision maker's opinion on what the probability is, and the objective probability which is derivable mathematically or as a result of observations. Thus, an optimist may consider the probability of being delayed in traffic very small and make a decision based on this assumption, while observation of traffic conditions over a period of time may indicate that such delays will occur 90% of the time. Probability and statistics play important roles in this effort. *See* STATISTICS.

Corresponding to each interaction of a strategy (S_i) and a state of nature (N_j), a consequence or outcome (O_{ij}) is determined in terms of some value or utility parameters. The outcome may sometimes be easily computed, for example, the interest cost on a business loan, but quite often must be based on insufficient data, past experience, or questionable assumptions, such as would be the case in trying to assess the outcome of a limited war. The generation of reliable outcome data can be expensive and time-consuming, and a probability of accuracy is often associated with each outcome. How to assess the relative value or utility of mixed parameters to present a single parameter for the outcome is still a paramount question in decision and utility or value theory, that is, establishing the utility of two apples versus one apple and one orange versus two oranges and so forth.

The information in the decision matrix is used to generate the last column which describes the overall expected outcome, expected value (EV), expected utility (EU), or expected profit for each strategy.

The decision tree of Fig. 1 and the decision matrix of Fig. 2 have been discussed assuming that the probabilities are known, such that these appear to represent the case 2, decisions under probabilistic uncertainty. Alternatively, if various probabilities can be assumed to be 0 or 1, a decision-under-certainty situation is obtained. If the probabilities are inexactly specified, then case 3, probabilistic imprecision, results. Finally, if the outcome values or utilities are specified inexactly, case 4 results.

Determination of utility. When choosing among alternatives, the decision maker must be able to indicate preferences among decisions that may result in diverse outcomes. In simple situations when only money is involved, an expected-value approach might be suggested, in which a larger expected amount of money is preferred to a smaller amount. However, in many situations the utility associated with money is not a linear function of the amount of money involved.

Risk aversion issues also need to be considered. One decision maker may have an aversion toward risk, while another may be risk-prone and make decisions that are associated with high-impact unfavorable consequences, perhaps because of very large rewards if the chance outcome does turn out to be favorable. Not all consequences are easily quantified in terms of monetary value, especially when there are multiple and potentially incommensurate objectives. A scalar utility function makes it possible, in principle, to attempt to resolve questions which involve uncertain outcomes with multiple attributes. The notion of a lottery provides a basis for the utility concept.

A utility function is a mathematical transformation which maps the set of outcomes of a decision problem into some interval of the real line, thereby assigning a numerical value to each outcome of a decision problem. It conforms with the basic set of four assumptions, conditions, or axioms needed to establish a utility theory. Axiom 1 is often called the orderability condition. It ensures that the preferences of the decision maker impose a complete ordering and a transitive preference ordering across outcomes. Axiom 2 is a monotonicity condition. It ensures that the preference function is smooth and that increasing preferences are associated with increasing rewards.

Axiom 3 is a continuity condition that ensures comparability across outcomes with equivalent preferences. Axiom 4 is a decomposition condition that ensures transformation of lotteries and substitutability of equivalent lotteries.

While utility theory is not complicated, successfully practicing the theory is very complicated. A number of important theoretical issues can easily be overlooked.

According to expected utility theory, the decision maker should seek to choose the alternative a_i which makes the resulting expected utility the largest possible. The utility u_{ij}, of choosing decision a_i and obtaining outcome event e_j, will also depend upon the particular value of the probabilistically uncertain random variable e_j as conditioned on the decision path that is selected. So, the best that the decision maker can do here is to maximize some function, such as the expected value of utility, as in Eq. (1), where

$$\max_i \text{EU}\{a_i\} = \max_i \sum_{j=1}^{n} u_{ij} P(e_j \mid a_i) \qquad (1)$$

the maximization is carried out over all alternative decisions, and $P(e_j \mid a_i)$ is the probability that the state of nature is e_j given that alternative a_i is implemented. The abbreviated notation $\text{EU}\{a_i\}$ is often used to mean the expected utility of taking action a_i. Generally, this is also called the subjective expected utility (SEU). "Subjective" denotes the fact that the probabilities may be based on subjective beliefs and the utilities may reflect personal consequences.

Generation of a utility curve. A lottery is a chance mechanism which results in an outcome with a prize or consequence $e(1)$ with probability $P(1)$, an outcome with prize or consequence $e(2)$ with probability $P(2), \ldots,$ and an outcome with prize or consequence $e(r)$ with probability $P(r)$. The probabilities $P(i)$ must be nonnegative and such that they sum to one, so that Eq. (2) holds. The lottery is denoted as $L = [e_1, P_1; e_2, P_2; \ldots; e_N, P_N]$.

$$\sum_{j=1}^{M} P(j) = P(1) + P(2) + \cdots + P(M) = 1 \qquad (2)$$

A utility curve for money can be generated in the following manner. All the outcomes of a decision are arranged in order of preference. The most preferred outcome is denoted W and the least preferred outcome L. The utilities $U(W) = 1$ and $U(L) = 0$ are arbitrarily assigned. Next, the decision maker is encouraged to answer the question: "Suppose I owned the rights to a lottery which pays W with probability 1/2 and L with probability 1/2. For what amount, say $X_{0.5}$, would I be willing to sell the rights to this lottery?" Since the decision maker is indifferent between definitely receiving $X_{0.5}$ and participating in the lottery $[W, 0.5; L, 0.5]$, Eq. (3) must be satisfied (**Fig. 3**).

$$U[X_{0.5}] = 0.5U(W) + 0.5U(L) = 0.5 \qquad (3)$$

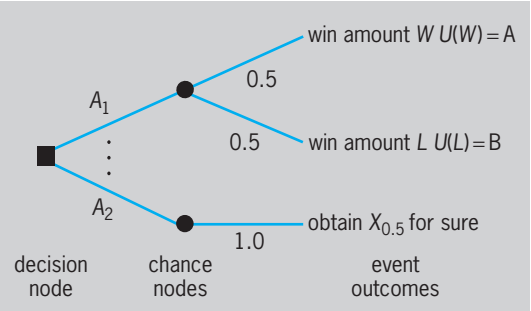


Fig. 3. Decision tree structure to determine the subjective utility of lottery outcomes.

Any other lottery with outcomes W and L could have been picked, such as $[W, 0.01; L, 0.99]$. There are several analytical reasons, however, for choosing a lottery which gives W or L with equal probability, or $[W, 0.5; L, 0.5]$. It may be conceptually quite simple for the decision maker to imagine a lottery which gives W or L with equal probability. In this case, the decision maker need only be concerned with the relative preferences of W and L. Also, the value of the lottery $[W, 0.5; L, 0.5]$ occurs midway on the U axis between W and L and therefore facilitates plotting the utility curve. This choice of 0.5 probability for the assessment may, however, be quite unrealistic. For example, W might mean some enormous increase in profit to the firm and L might mean going bankrupt. A business person might so abhor a situation in which there is a 50% chance of going bankrupt that it would not be realistically possible to think in terms of such extreme utilities.

Next, the decision maker is encouraged to answer the question: "Suppose I owned the rights to a lottery which paid W with probability 0.5 and $X_{0.5}$, which has a utility of 0.5, with probability 0.5. For what amount, call it $X_{0.75}$, would I just be willing to sell the rights to this lottery?" Here, the decision maker is indifferent between receiving $X_{0.75}$ for certain and participating in the lottery $[W, 0.5; X_{0.5}, 0.5]$, so Eq. (4) holds. Next, the lottery $[X_{0.5}, 0.5; L, 0.5]$

$$U(X_{0.75}) = 0.5U(X_{0.5}) + 0.5U(L) = 0.75 \qquad (4)$$

is considered, and the amount to be determined is $X_{0.25}$. In this case, Eq. (5) is obtained. This process

$$U(X_{0.25}) = 0.5U(X_{0.5}) + 0.5U(L) = 0.25 \qquad (5)$$

could be continued indefinitely. However, the three points generated here and the two end points are often sufficient to give an idea of the shape of the utility curve, which is usually concave downward, concave upword, or flat (**Fig. 4**). The utilities that have been obtained are indicators of preference, and are not absolute measurements. As a result, they are unique only up to a linear transformation of the form $U'(X) = aU(X) + b$, where $a > 0$.

A decision maker is rarely risk-prone, and most people are risk-averse in almost all situations. One of the few exceptions to this behavior is situations where there is a very small probability of receiving a very large reward, such as a probability of 10^{-6} of

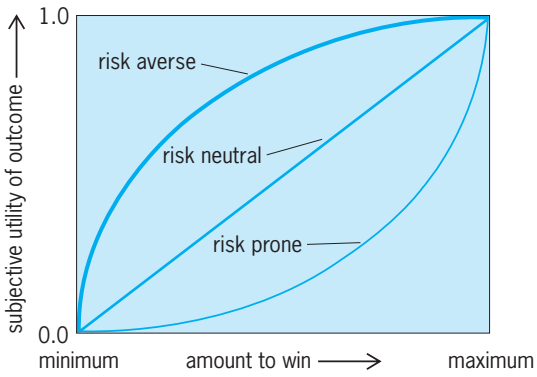


Fig. 4. Hypothetical subjective utility curves.

receiving 10^6, and a person is willing to pay \$5 for such a lottery. Thus, the typical utility curve is concave downward. There are other notable exceptions to this generally risk-averse nature of the decision maker, especially when the utility function is not required to be symmetric about losses and gains.

Utility assessment. The first step in a utility assessment process ideally results in a definition of the scope of the assessment procedure. Elements of the value system are identified and often structured, at least partially, in the issue formulation step. Attributes, according to which alternatives are to be assessed, are identified and structured. The attributes are next related to measurable characteristics of the alternatives. Much of this effort is accomplished in the formulation and analysis steps, and the results should be organized to match the requirements of the utility assessment portion of the interpretation step. The final part of the utility assessment process concerns either elicitation of the weights of each attribute and aggregation of the information to facilitate selecting the best alternative, or the use of a number of holistic judgments and regression analysis to identify these weights. *See* SYSTEMS ENGINEERING.

Utility functions are fully useful only when they are intelligently obtained and thoughtfully applied. Structuring of a decision assessment issue is a crucial part of an overall decision assessment process. The final result will, in practice, depend upon the truthfulness of both the decision situation structural model and the parameters within this structure.

Once the decision maker's utility curve has been defined, the standard tenets of axiomatic decision theory suggest that it may be employed in any decision problem as long as all the outcomes fall between the extremes used to establish the utility function. In principle, the decision maker can now delegate authority to an agent, who can employ this utility curve to arrive at the same decisions that the decision maker would make. However, in reality a person's utility curves continually change, as the person learns new things, has new experiences, or develops new perspectives. Similarly, a utility curve that applies to one kind of problem may offer little insight into another. Thus, in many situations it may not be appropriate to suggest that someone else can learn the decision maker's utility curves, and then operate as if these could be rigidly applied in changing circumstances.

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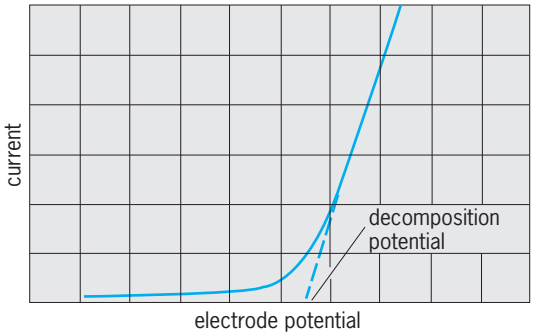
Decomposition potential

The electrode potential at which the electrolysis current begins to increase appreciably. Decomposition potentials are used as an approximate characteristic of industrial electrode processes. *See* ELECTROCHEMICAL PROCESS; ELECTROLYSIS.

Decomposition potentials are obtained by extrapolation of current-potential curves (see **illus.**).



Determination of decomposition potential.

Extrapolation is not precise because there is a progressive increase of current as the electrode potential is varied. The decomposition potential, for a given element, depends on the range of currents being considered. The cell voltage at which electrolysis becomes appreciable is approximately equal to the algebraic sum of the decomposition potentials of the reactions at the two electrodes and the ohmic drop, or voltage drop, in the electrolytic cell. The ohmic drop term is quite negligible for electrolytes with high conductance. For a discussion of current-potential curves. *See* ELECTROLYTIC CONDUCTANCE; OVERVOLTAGE.

Paul Delahay

Decompression illness

Symptoms in humans which result from a sudden reduction in atmospheric pressure. It is also called dysbarism, caisson disease, the bends, and compressed-air illness. It is most commonly seen in two groups

of subjects: those who rapidly ascend in nonpressurized airplanes to altitudes in excess of 18,000 ft (5500 m); and divers, scuba divers, sandhogs, and professional workers in hyperbaric chambers who work under increased ambient pressures and are decompressed. *See* DIVING.

The onset of symptoms may occur at any time from a few minutes to several hours after decompression. The most common manifestation is pain in the joints and muscles. However, skin, respiratory, and neurologic symptoms are not uncommon. The skin manifestations are itching, discoloration, and edema (swelling). Respiratory symptoms are coughing and dyspnea (difficulty in breathing). The neurologic symptoms are of a more grave nature and vary from mild paresthesia (sensations of tingling, crawling, or burning of the skin) and weakness to total paralysis; loss of bladder and rectal sphincter control is common. The severe forms of this illness are followed by circulatory failure, paralysis, coma, and death.

This condition is caused by the formation of nitrogen bubbles in the tissues and blood vessels. In the body the nitrogen normally dissolved in body fluids forms bubbles when the atmospheric pressure is reduced. These bubbles plug vessels and expand in tissue spaces, such as muscles and joints, producing the characteristic symptoms and signs of this illness. Helium, when present in high concentrations utilized by divers or found in some spacecraft, can also produce a similar condition.

Treatment is recompression, followed by gradual decompression to normal atmospheric pressure. Prognosis is generally good except in those subjects who show central nervous system damage. *See* AEROSPACE MEDICINE; BIOPHYSICS; SPACE BIOLOGY.

McChesney Goodall

Decontamination of radioactive materials

The removal of radioactive contamination which is deposited on surfaces or spread throughout a work area. Personnel decontamination is included. Radioactive contamination is a potential health hazard and, in addition, may interfere with the normal operation of plants, particularly when radiation detection instruments are used for control purposes. Thus, the purpose of decontamination is the detection and removal of radioactive contaminants from unwanted locations to locations where they do not create a health hazard or interfere with plant operation. The objective of a good decontamination operation is to remove the radioactive contamination permanently from the human environment, with minimum radiation exposure.

There are four ways in which radioactive contaminants adhere to surfaces, and these limit the decontamination procedures which are applicable. The contaminant may be (1) held more or less loosely by electrostatic forces or surface tension, (2) absorbed in porous materials, (3) mechanically bonded to surfaces through oil, grease, tars, paint, and so on, or (4) adsorbed or chemically bonded to the surface. Typically, the last of these provides the most adherent deposits, often requiring chemical decontamination methods.

Methods. Decontamination methods are mechanical or chemical. Commonly used mechanical methods are vacuum cleaning, sand blasting, blasting with solid carbon dioxide, flame cleaning, scraping, ultrasonic cleaning, vibratory finishing, and using lasers to vaporize contaminants. Scraping includes surface removal—for example, removal of a contaminated layer from concrete floors with an air hammer or scrabbling device, which scrapes off the top layer of concrete mechanically. Strippable coatings are sprayed onto a painted, metal, or concrete surface; contamination is removed when the dried coating is peeled off. Hydrolyzing with high-pressure water is another effective technique. The interior surfaces of piping systems can be cleaned out using rotating brushes (wire or plastic with abrasive tips) or water- or air-propelled abrasive devices.

Chemical methods are used primarily to decontaminate components and tools that are immersed in a tank, either by means of a chemical solvent to dissolve the contaminant, or by using electropolishing techniques to remove the surface layer, including contaminants, from metals. Chemical decontamination methods are also used to remove radioactive deposits from the interior surfaces of piping, pumps, heat exchangers, and boilers. For these applications, the solvent is pumped or flushed through the system, dissolving the radioactive deposits. The solution itself is then radioactive, and the contaminants are typically removed using filters or ion-exchange resins. The use of this approach to clean the coolant systems of nuclear reactors has become common. Dilute chemical reagents, including organic acids, are used to decontaminate the primary coolant systems of operating nuclear power plants to minimize radiation exposure of the workers. An important consideration for plants that are returning to service is to avoid any corrosion of the system components; in these cases, typically 90–95% of the radioactivity is removed, thereby reducing radiation levels by a factor of 10–20. For permanently shutdown plants, more aggressive solvents can be used; complete removal of contaminants is possible, permitting recycling of the metal components.

Collection and disposal of the radioactive residue from mechanical decontaminations can be more difficult, and this is often a major consideration in determining which process to use. In some cases, dust is produced, which has to be removed from the atmosphere using high-efficiency filters. The wet processes, such as hydrolyzing, result in large quantities of water, which then has to be treated. In all cases, it is important that the volume of the final radioactive waste be as small as possible, because large volumes cost more to store and transport to a repository, and disposal charges are greater.

Other considerations, besides waste treatment and disposal, that influence the choice of decontamination method include (1) cost, (2) hazards involved

in the procedure, (3) impact on other plant systems and operations (for example, the need to isolate areas for dust and debris removal), and (4) permanency of the removal (for example, some processes may leave the surface in a rougher condition, resulting in rapid recontamination). For objects that are contaminated with short-lived radioactive material, such as fission products, it may be more cost-effective to postpone the decontamination to allow for radioactive decay. In cases with alpha-emitting contamination, painting of the surface to fix the contaminant in place may provide a permanent solution. *See* ALPHA PARTICLES.

Personnel. Personnel decontamination methods differ from those used for materials primarily because of the possibilities of injury to the subject. Procedures used for normal personal cleanliness usually will remove radioactive contaminants from the skin, and the method used will depend upon the form of the contaminant and associated dirt (grease, oil, soil, and so on). Soap and water (sequestrants and detergents) normally remove more than 99% of the contaminants. If it is necessary to remove the remainder, chemical methods which remove the outer layers of skin can be used. These chemicals, such as citric acid, potassium permanganate, and sodium bisulfite, should be used with caution and preferably under medical supervision because of the risk of injury to the skin surface. The use of coarse cleansing powders should be avoided for skin decontamination, because they may lead to scratches and abraded skin which can permit the radioactive material to enter the body. The use of organic solvents should be avoided because of the probability of penetration through the pores of the skin. It is very difficult to remove radioactive material once it is fixed inside the body, and the safest and most reliable way to prevent radiation exposure from contamination is the application of health physics procedures to prevent entry of radioactive material into the body.

Environment. Air contaminants frequently are eliminated by dispersion into the atmosphere. Certain meteorological conditions, such as prevailing wind velocity and direction, and inversion layers, seriously limit the total amount of radioactive material that may be released safely to the environment in any time period. Consequently, decontamination of the airstream by high-efficiency filters, cyclone separators, scrubbing with caustic solutions, cryogenic removal, and entrapment on charcoal beds is often implemented. The choice of method is guided by airflow requirements, the cost of heating and air conditioning, the cost of waste storage and disposal, the hazards associated with the airborne radioactive material, and the isolation of the operation from other populated areas. *See* AIR FILTER; DUST AND MIST COLLECTION; FILTRATION.

Water decontamination processes can use either maximum dilution or maximum concentration (and subsequent removal) of the contaminant, using water purification processes such as ion exchange, chemical precipitation, flocculation, filtration, and evaporation. Contaminated soil is typically cleaned by washing with dilute chemical solutions of chelating agents, such as ethylenediaminetetraacetic acid (EDTA), or is removed to a regulated landfill. *See* ION EXCHANGE; WATER TREATMENT.

Certain phases of radioactive decontamination procedures are potentially hazardous to personnel. Health physics practices include the use of protective clothing and respirators, localized shielding, isolation or restriction of an area, and the use of remotely controlled equipment. *See* HAZARDOUS WASTE; HEALTH PHYSICS; RADIATION INJURY (BIOLOGY); RADIOACTIVE WASTE MANAGEMENT; RADIOACTIVITY.

Christopher J. Wood

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Deep inelastic collisions

Either highly energetic collisions of elementary particles, namely, leptons and nucleons, which probe the nucleons' internal structure; or collisions between two heavy ions in which the two nuclei interact strongly while their nuclear surfaces overlap.

Elementary-particle collisions. Many collisions between elementary particles either are elastic, in which the particles emerge unscathed, or proceed via the creation of a resonance, in which the particles briefly coalesce and the resonance subsequently decays. Deep inelastic collisions of elementary particles are very energetic collisions between leptons such as electrons or neutrinos and nucleons (that is, protons or neutrons, typically in a nucleus) in which the target nucleon breaks up into many particles and the lepton is scattered through a large angle in the center-of-mass frame. These collisions are akin to the Rutherford scattering experiments in which most alpha particles went through a thin gold foil undeflected but some were deflected through large angles. In both cases, the explanation for large deflections is that the incident particle encounters not a uniform sphere of material but a few hard or pointlike objects inside the target. The alpha particles encounter gold nuclei, while leptons strike quarks inside the nucleons. *See* ALPHA PARTICLES; LEPTON; NUCLEON; RESONANCE (QUANTUM MECHANICS); SCATTERING EXPERIMENTS (NUCLEI).

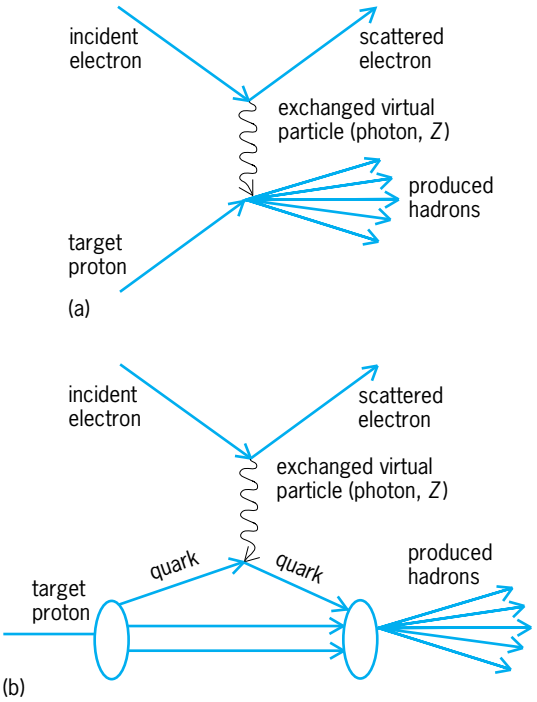
Discovery of quarks. Quarks were originally postulated by M. Gell-Mann and others in the early 1960s mainly as a mathematical device to explain the observed symmetries in elementary particles known at the time. Deep inelastic scattering provided the first hard evidence of their actual existence inside protons and neutrons, giving quarks a new measure of reality.

In a deep inelastic collision of a lepton of energy E with a nucleon of mass M, the scattering is usually described as a function of two variables: Q^2, which may be thought of as the square of the energy of the exchanged virtual particle, and y, which is the fraction of energy lost by the lepton. In the late 1960s J. Bjorken predicted the onset of scaling at very high energy transfers, that is, the scattering would become a function not of Q^2 and y independently but of the ratio $x \equiv Q^2/(2MEy)$ alone. This was dramatically confirmed in a series of historic electron scattering experiments conducted around 1968. R. Feynman soon showed that an interpretation of the observed collisions in terms of elastic scattering of leptons from pointlike parts of nucleons, which he called partons, leads directly to the same Bjorken scaling. The variable x is simply the fraction of the nucleon's momentum carried by the parton. With these experiments and Feynman's interpretation came the realization that protons and neutrons are not fundamental but are in turn composed of smaller objects that are now called quarks. A whole new substructure of matter had been discovered. *See* QUARKS.

Structure functions. Deep inelastic scattering experiments are now conducted to study the structure of protons and neutrons. In each collision the fraction x of the nucleon's momentum carried by the struck quark is measured, and thus the x distributions of quarks inside a proton are directly measured. These are known as structure functions. Studies of these have shown that the momentum of a proton is not carried entirely by quarks. In fact, only about half the momentum can be ascribed to quarks. The other half is believed to be carried by gluons, which are carriers of the strong force which binds the quarks within nucleons and other hadrons. By studying the evolution of quark structure functions as Q^2 changes, the structure functions of gluons may be inferred by using the theory of strong interactions called quantum chromodynamics (QCD). *See* GLUONS; QUANTUM CHROMODYNAMICS.

The quark structure functions may also be used to deduce the number of valence quarks inside a nucleon. A valence quark, unlike a sea quark, does not arise from a gluon creating a quark-antiquark pair and hence does not have a partner antiquark. Thus, when the number of antiquarks is subtracted from the number of quarks, only the number of valence quarks remains. This relation is called the Gross-Llewellyn-Smith sum rule, and experiments confirm that there are three valence quarks in a nucleon as expected in the quark model of hadrons. Similarly, other sum rules are tested by using structure functions measured by deep inelastic scattering experiments. Attention has focused on measuring the structure functions for quarks with spin aligned or antialigned to the direction of the proton's spin. The surprising conclusion is that only a small fraction of the spin is carried by valence quarks. *See* SUM RULES.

Deep inelastic scattering may also be viewed as a collision between the virtual particle exchanged (see **illus.**) and the nucleon. When viewed in this fashion,



Deep inelastic collision between an electron and a proton where the nucleon is viewed either (*a*) as a blob that breaks up or (*b*) as composed of quarks, one of which is ejected by elastic scattering with the electron.

it is possible to measure the ratio of the scattering cross section from (hypothetical) spin-zero partons and spin-half partons by using the measured cross section for longitudinally and transversely polarized virtual probes. This ratio is expected to be zero and is indeed measured to be close to zero. The difference is understood in QCD as being due to the effect of gluon radiation.

Experiments. Modern-day experiments in deep inelastic scattering often use neutrinos or muons as probes. Neutrinos have the advantage that they are not affected by the electric charge of the target nucleus and hence scatter directly off the quarks, making the study of structure functions with neutrinos particularly attractive. Muons are easy to detect and identify. However, the highest-energy deep inelastic collisions are carried out by using an electron-proton collider. These experiments aim to study structure functions at very low values of x where some models predict new behavior.

It is natural to ask whether there exists yet another layer of tiny particles, smaller than even the quarks. Since quarks cannot be isolated for such a study, the highest-energy deep inelastic collisions use electrons as probes to search for quark substructure.

Milind V. Purohit

Heavy-ion collisions. Deep inelastic collisions of heavy ions are characterized by features that are intermediate between those of comparatively simple quasielastic, few-nucleon transfer reactions and those of the highly complex compound-nucleus reactions. These deep inelastic or damped collisions often occur for heavy-ion reactions at center-of-mass energies less than 5 MeV per nucleon above the

Coulomb barrier. During the brief encounter of the two nuclei, large amounts of kinetic energy of radial and orbital motion can be dissipated. The lifetime of the dinuclear complex (analogous to a chemical molecule) corresponds to the time required for the intermediate system to make a partial rotation (10^{-22} s to 5×10^{-21} s). On separation, the final total kinetic energies of the two reaction fragments can be well below those corresponding to the Coulomb repulsion of spheres, indicating that the fragments are highly deformed in the exit channel, as is known to be the case for fission fragments. *See* NUCLEAR FISSION.

Characteristic features. The principal properties of damped or deep inelastic reactions at bombarding energies of a few megaelectronvolts per nucleon above the interaction barrier are well established. Among these characteristic features are the following: (1) The reactions are binary with two massive reaction fragments in the exit channel. Subsequently, the excited primary fragments, which for large-energy losses have nearly the same temperature, decay via fission or the emission of light particles and gamma rays. (2) Large amounts of kinetic energy of radial and orbital motion can be dissipated in a reaction. (3) Fragment mass and charge distributions are bimodal and are centered close to the masses and charges of projectile and target nuclei. The second moments of these distributions are strongly correlated with kinetic energy dissipation. However, the first moment in $A(Z)$ is nearly independent of energy loss. (4) The angular distributions of the reaction fragments change in a characteristic way with the charge product $Z_P \cdot Z_T$ of the projectile-target combination and the bombarding energy.

Reaction fragment distributions. Experimental studies of the first and second moments of the mass, charge, and isobaric and isotopic distributions of damped reaction fragments and the variation of the moments with total kinetic energy loss and scattering angle have yielded new insight into the microscopic reaction mechanism for heavy ions. Evidence for the magnitude of the charge and mass flow is manifested by the second moments of these distributions which increase with increasing energy loss (where energy loss is a direct measure of the reaction time). Studies of the second moment (variance) in the charge distribution for fixed mass asymmetry (isobaric distribution) have shown that the charge equilibration degree of freedom, like the excitation energy, is relaxed in less than 2×10^{-22} s. In contrast to the short relaxation times for the above two intrinsic degrees of freedom, the mass asymmetry degree of freedom relaxes far more slowly.

Light-particle emission. Studies of light-particle emission from deep inelastic reactions are important because these particles can, in principle, be emitted at any stage of the reaction and hence carry signatures of the dominating mechanisms at work at the different reaction stages. In damped reactions with bombarding energies up to 5 MeV per nucleon above the Coulomb barrier, where hundreds of megaelectronvolts of energy can be dissipated, neutron spectra are consistent with evaporation from the two fully accelerated heavy fragments. For kinetic energy losses greater than 50 MeV, analysis of the neutron spectra indicates that the two fragments have essentially the same temperature as a function of energy loss. In addition, analysis of the differential cross sections of emitted neutrons leads to the conclusion that the excitation energy is divided between the fragments in proportion to their masses. These interesting and important results are, at first, somewhat puzzling when interpreted in terms of simple nucleon exchange models, where the lighter fragment is predicted to have the higher temperature. However, such an initial temperature gradient over the dinuclear complex can be expected to decay through subsequent interactions. The nucleon exchange mechanism itself serves to produce a strong temperature-equalizing effect over the two components of the dinuclear complex with a temperature gradient. The dynamical competition between generation of a temperature difference and its decay can only be estimated from realistic model calculations.

Interpretation. The ultimate goal of experimental and theoretical investigations of damped or deep inelastic reactions between complex nuclei is to achieve an understanding of the reaction properties in terms of the underlying microscopic mechanisms. However, the development of microscopic quantal reaction theories has not progressed far enough to yield a consistent explanation of the observed reaction features. In this situation, a phenomenological interpretation appears to be useful, where certain reaction patterns are analyzed in terms of macroscopic or microscopic concepts, making it possible to study the dependence of model parameters on important reaction conditions. Comparisons of a variety of experimental results with dynamical transport calculations based on one-body nucleon exchange give strong evidence that stochastic exchange of independent nucleons accounts for the dominant part of the dissipative and fluctuative phenomena observed in damped reactions. However, the exact fraction of the energy loss that is due to nucleon exchange and the fraction due to other, more minor processes, including collective modes of energy transformation, remain to be determined. *See* NUCLEAR REACTION; NUCLEAR STRUCTURE.

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Deep-marine sediments

The term "deep marine" refers to bathyal sedimentary environments occurring in water deeper than 200 m (650 ft), seaward of the continental shelf break, on the continental slope and the basin (**Fig. 1**). The continental rise, which represents that part of the continental margin between continental

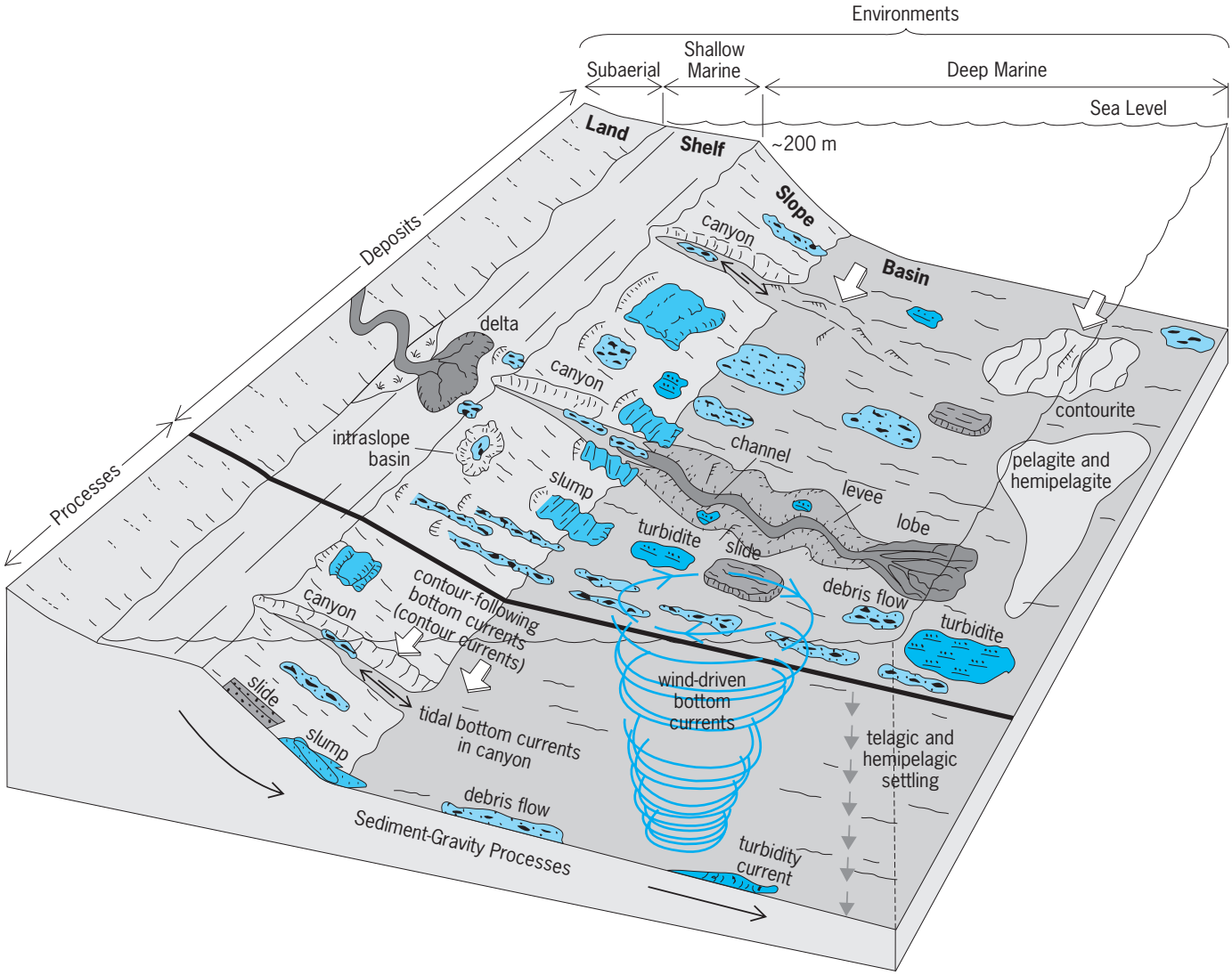


Fig. 1. Slope and basinal deep-marine sedimentary environments occurring at water depths greater than 200 m (650 ft). Slides, slumps, debris flows, turbidity currents, and various bottom currents are important processes in transporting and depositing sediment in the deep sea. Note the complex distribution of deep-marine deposits.

slope and abyssal plain, is included under the broad term "basin." An example of well-developed shelf, slope, and basin settings can be seen in the modern Gulf of Mexico (**Fig. 2**). On the slope and basin environments, sediment-gravity processes (slides, slumps, debris flows, and turbidity currents) and bottom currents are the dominant depositional mechanisms, although pelagic and hemipelagic deposition is also important. *See* BASIN; CONTINENTAL MARGIN; GULF OF MEXICO; MARINE SEDIMENTS.

In 1960s and 1970s, most deep-marine sedimentary systems were thought to be dominated by submarine-fan deposits formed by turbidity currents regardless of whether they were fan-shaped in morphology or whether they were turbidites in origin. A critical evaluation of deep-marine sedimentary systems revealed that, in fact, they are quite complex and that simple fan models are not realistic. Furthermore, the basic tenets of understanding deep-marine processes and their deposits have recently been challenged. In light of these developments, this article focuses on types of deep-marine processes and their deposits; and discusses aspects of deep-marine sedimentary environments in the categories of (1) submarine slopes, (2) submarine canyons and gullies, (3) submarine channels, (4) submarine fans, and (5) submarine basin plains. *See* SUBMARINE CANYON; TURBIDITE; TURBIDITY CURRENT.

Types of processes. The mechanics of deep-marine processes is critical in understanding the nature of transport and deposition of sand and mud in the deep sea. In deep-marine environments, gravity plays the most important role in transporting and depositing sediments. Sediment failure under gravity near the shelf edge commonly initiates gravity-driven deep-marine processes, such as slides, slumps, debris flows, and turbidity currents (Fig. 1). Sedimentary deposits reflect only depositional mechanisms, not transportational mechanisms. Therefore, general characteristics of each process listed below should be used only to interpret depositional mechanisms. *See* GRAVITY; SEDIMENTOLOGY.

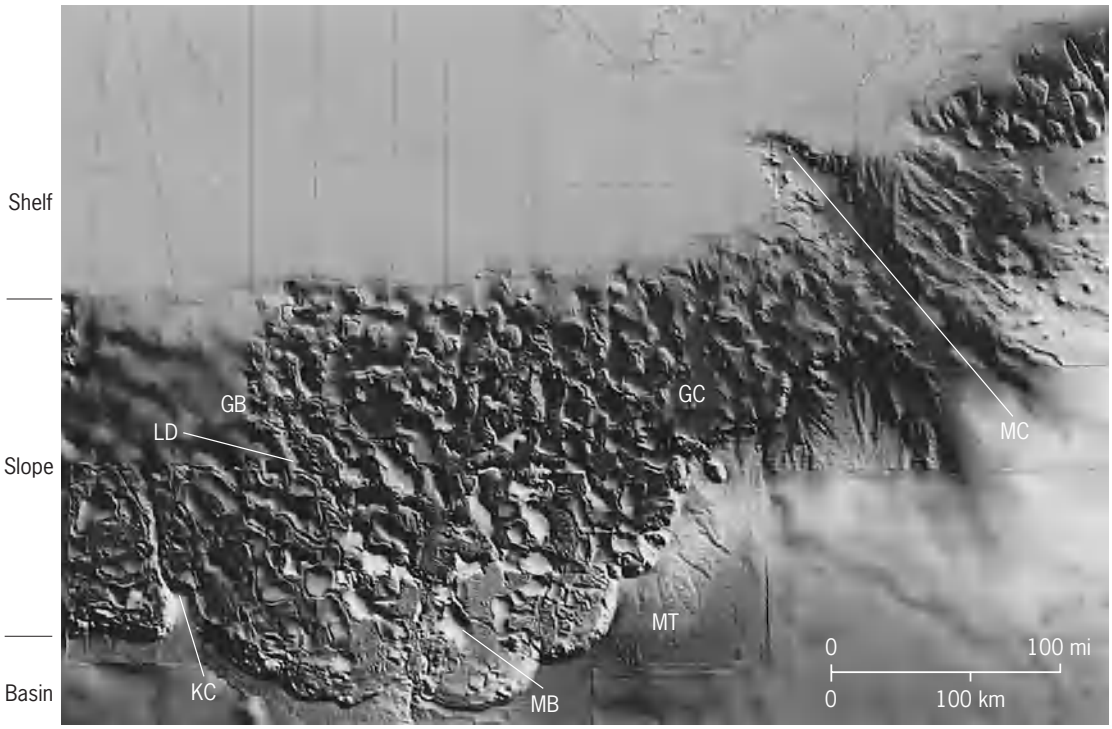


Fig. 2. Bathymetric display of the northern Gulf of Mexico based on high-resolution bathymetric swath data. The data are displayed to give a 3-D perspective of present-day shelf, slope, and basin settings. Note the highly irregular sea-floor topography and intraslope basins controlled by salt tectonism. The intraslope basins commonly range in width from 10 to 20 km (6.2 to 12.4 mi).

Slides. A slide is a coherent mass of sediment that moves along a planar glide plane and shows no internal deformation (Fig. 1). Slides represent translational movement. Submarine slides can travel hundreds of kilometers. For example, the runout distance of Nuuanu Slide in offshore Hawaii is 230 km (143 mi). Long runout distances of 50–100 km (31–62 mi) of slides are common.

General characteristics of submarine slide deposits are:

- Gravel-to-mud lithofacies
- Large dimensions, up to hundreds of kilometers long
- Glide plane scar on the main shear surface, and basal zone of shearing
- Upslope areas contain tensional faults
- Downslope edges may show compressional folding or thrusting (that is, toe thrusts)
- Common in areas of tectonics, earthquakes, salt movements, and rapid sedimentation
- Sheetlike geometry with irregular thickness and shape
- Occur on slopes as gentle as 1–4°
- Travel hundreds of kilometers down the gentle slopes (**Fig. 3**)

Examples of submarine slides reveal that modern ones are an order of magnitude larger than their known ancient counterparts. The modern Storegga Slide on the Norwegian continental margin has a width of 290 km (180 mi), a thickness of 430 m (1411 ft), and a width-to-thickness ratio of 675:1 (see **table**). A seismic-scale slide sheet was described, 6 km (3.7 mi) wide and 440 m (1444 ft) thick with a width-to-thickness ratio of 14:1, from the Ablation Point Formation (Kimmeridgian) exposed in Alexander Island of Antarctica. Individual sandy slide sheets reach a thickness of 50 m (164 ft) and a length of 1000 m (3281 ft).

Slumps. A slump is a coherent mass of sediment that moves on a concave-up glide plane and undergoes rotational movements causing internal deformation (Fig. 1).

General characteristics of submarine slump deposits are:

- Gravel-to-mud lithofacies
- Basal zone of shearing
- Slump folds
- Contorted bedding
- Steeply dipping and truncated layers
- Undeformed units above deformed units
- Lenticular-to-sheetlike geometry with irregular thickness

For example, one study reported a thickness of slumps varying from 10 to 374 m (33 to 1227 ft).

Lateral dimensions (width/length) reached up to 64 km (40 mi). Width-to-thickness ratio varied from 10:1 to 3600:1 (see table). In the Cretaceous of the Norwegian North Sea, abrupt lateral changes in slump facies thickness occur across a distance of 29 m (95 ft) between two adjacent wells. Such abrupt terminations of slump deposits have been observed in outcrops in Antarctica. Slump and slide deposits show large vertical dimensions. *See* ANTARCTICA; CRETACEOUS.

Debris flows. A downslope increase in mass disaggregation results in the transformation of slumps into debris flows (Fig. 1). Sediment is now transported as an incoherent viscous mass, as opposed to a coherent mass in slides and slumps. A debris flow is a sediment-gravity flow with plastic rheology (that is, fluids with yield strength) and laminar state. Deposition from debris flows occurs through freezing. The term "debris flow" is used here for both the process and the deposit of that process. The terms "debris flow" and "mass flow" are used interchangeably because each exhibits plastic flow behavior with shear stress distributed throughout the mass. Although only muddy debris flows (debris flows with mud matrix) received attention in the past, recent experimental and field studies show that sandy debris (debris flows with sand matrix) flows are equally important. Rheology is more important than grain-size distribution in controlling sandy debris flows, and the flows can develop in slurries of any grain size (very fine sand to gravel), any sorting (poor to well), any clay content (low to high), and any modality (unimodal and bimodal). *See* RHEOLOGY.

General characteristics of debris-flow deposits (both muddy and sandy types) are:

- Gravel-to-mud lithofacies
- Floating or rafted mudstone clasts near the tops of beds
- Inverse grading of floating mudstone clasts
- Inverse grading, normal grading, inverse-to-normal grading, and no grading of matrix
- Floating quartz pebbles and inverse grading of granules
- Pockets of gravels
- Planar clast fabric
- Preservation of fragile shale clasts
- Irregular, sharp upper contacts and lateral pinch-out geometries
- Side-by-side occurrence of garnet granules (density 3.5–4.3) and quartz granules (density 2.65)
- Lenticular-to-sheetlike geometry

Using outcrop measurements, it can be established that debris-flow deposits are likely to have a width-to-thickness ratio of 30–50:1. In the examples used in this study, thicknesses of debris-flow deposits commonly range from 1 to 60 m (3 to 197 ft), but unusually thick deposits may occur in association with large slides. An example is the Storegga Slide in the Norwegian continental margin, 430 m (1411 ft) thick and 290 km (180 mi) wide, in which debris flow is a dominant facies. Debris-flow facies have width-to-thickness ratios of up to 500:1 or more. This is generally the result of amalgamation of depositional units.

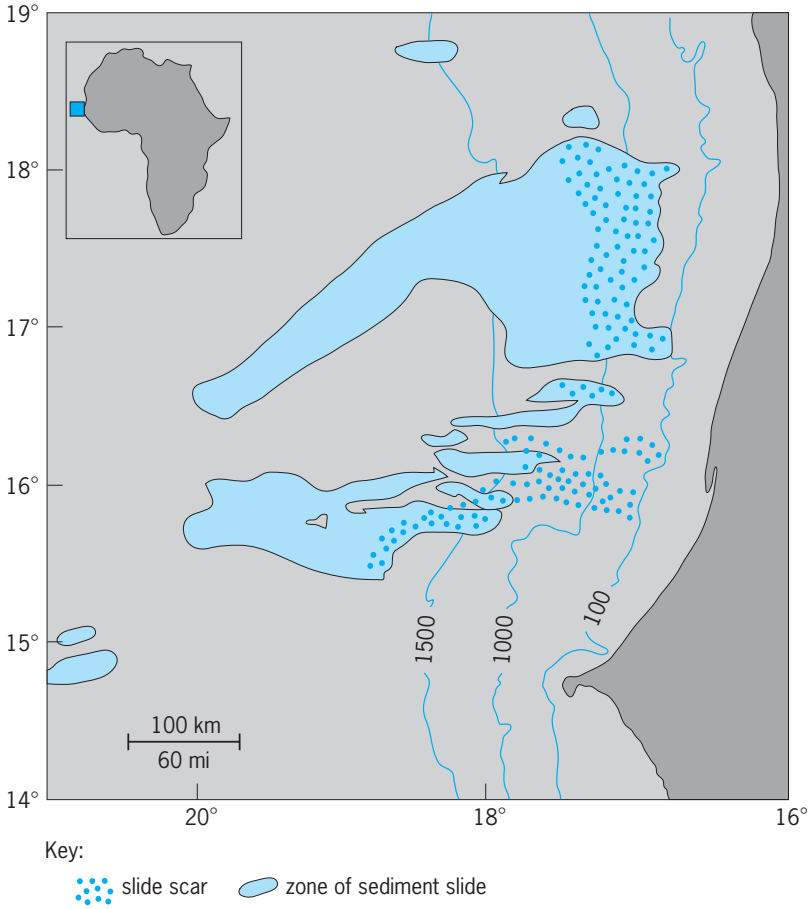


Fig. 3. Map showing long-distance transport of modern slides of about 300 km (186 mi) downslope from the shelf edge (100-fathom, 183-m, or 600-ft contour line) off northwestern Africa. Thickness of slide = 25 m (82 ft). (*From R. D. Jacobi, Sediment slides on the northwestern continental margin of Africa, Mar. Geol., 22:157–173, 1976*)

Turbidity currents. With increasing fluid content, plastic debris flows tend to become newtonian turbidity currents (Fig. 1). Although turbidity currents can constitute a distal end member occurring in basinal areas, they can occur in any part of the system (proximal and distal). Turbidity currents can also occur above debris flows due to flow transformation in density-stratified flows. A turbidity current is a sediment-gravity flow with newtonian rheology (that is, fluids without yield strength) and turbulent state. Deposition from turbidity currents occurs through suspension settling. Deposits of turbidity currents are called turbidites. Although turbidity currents have received a lot of emphasis in the past, other processes are equally important in the deep sea (Fig. 1). In terms of transporting coarse-grained sediment into the deep sea, sandy debris flows and other mass flows appear to play a greater role than turbidity currents.

General characteristics of turbidites are:

- Fine-grained sand-to-mud lithofacies
- Sharp or erosional basal contact
- Gradational upper contact

Dimensions of deep-marine facies

Example	Width:thickness ratio
Slide	
Lower Carboniferous, England	7:1 (100 m wide/long, 15 m thick)
Cambrian-Ordovician, Nevada	30:1 (30 m wide/long, 1 m thick)
Jurassic, Antarctica	45:1 (20 km wide/long, 440 m thick)
Modern, U.S. Atlantic margin	40–80:1 (2–4 km wide/long, 50 m thick)
Modern, Gulf of Alaska	130:1 (15 km wide/long, 115 m thick)
Middle Pliocene, Gulf of Mexico	250:1 (150 km wide/long, 600 m thick)
Slide/slump/debris flow/turbidite	
5000–8000 years B.P., Norwegian continental margin	675:1 (290 km wide/long, 430 m thick)
Slump	
Cambrian-Ordovician, Nevada	10:1 (100 m wide/long, 10 m thick)
Aptian-Albian, Antarctica	10:1 (3.5 km wide/long, 350 m thick)
Slump/slide/debris flow	
Lower Eocene, Gryphon Field, U.K.	20:1 (2.6 km wide/long, 120 m thick)
Paleocene, Faeroe Basin, north of Shetland Islands	28:1 (7 km wide/long, 245 m thick)
Slump	
Modern, southeast Africa	170:1 (64 km wide/long, 374 m thick)
Carboniferous, England	500:1 (5 km wide/long, 10 m thick)
Lower Eocene, Spain	900–3600:1 (18 km wide/long, 5–20 m thick)
Debris flow	
Modern, British Columbia	12:1 (50 m wide/long, 4 m thick)
Cambrian-Ordovician, Nevada	30:1 (300 m wide/long, 10 m thick)
Modern, U.S. Atlantic margin	500–5000:1 (10–100 km wide/long, 20 m thick)
Quaternary, Baffin Bay	1250:1 (75 km wide/long, 60 m thick)
Turbidites (depositional lobes)	
Cretaceous, California	165:1 (10 km wide/long, 60 m thick)
Lower Pliocene, Italy	1200:1 (30 km wide/long, 25 m thick)
Turbidites (basin plain)	
Miocene, Italy	11,400:1 (57 km wide/long, 5 m thick)
16,000 years B.P., Hatteras Abyssal Plain	125,000:1 (500 km wide/long, 4 m thick)

Normal grading without complications (that is, no floating clasts or granules)
Sheetlike geometry in basinal settings. Lenticular geometry may develop in channel settings

In the geologic record, units showing normal grading commonly occur directly above units with inverse grading. Such inverse–to–normally graded beds have traditionally been interpreted as high-density turbidites. However, the concept of high-density turbidity currents and their deposits have come under severe criticism because there are at least five different ways that the concept of high-density turbidity currents has been defined: (1) flow density, (2) grain size, (3) driving force, (4) velocity, and (5) rate of deposition. These five concepts are not in agreement in terms of fluid dynamics. As a result, there are no standard and meaningful ways of interpreting deep-water rocks as high-density turbidites. In terms of fluid rheology and flow state, high-density turbidity currents are nothing but sandy debris flows. *See* FLUID MECHANICS; GEOLOGIC TIME SCALE.

Submarine fan lobes (lobate deposits that accumulate at the mouths of canyons and channels) show width-to-thickness ratios of 165:1 to 1200:1. Lateral dimensions of turbidites deposited on modern abyssal plains can be unusually large. For example, the "Black Shell turbidite" on the Hatteras Abyssal Plain in the western North Atlantic Ocean is 4 m (13 ft) thick and extends for 500 km (311 mi), with a width-to-thickness ratio of 125,000:1. It covers an area of 44,000 km^2 (16,984 mi^2). Classic examples of basin-plain turbidites, such as those exposed along the foreshore at Zumaya, Spain, are usually in the range of 10 cm to 1 m (4 in. to 3 ft) in thickness, but they can be traced for several kilometers. Extensive sheetlike turbidites deposited in the open oceans of Atlantic-type margins are seldom preserved in the geologic record.

Bottom currents. In large modern ocean basins, such as the Atlantic, thermohaline-induced geostrophic bottom currents within the deep and bottom water masses commonly flow approximately parallel to bathymetric contours (that is, along the slope; Fig. 1). They are generally referred to as contour currents. However, because not all bottom currents follow regional bathymetric contours, it is preferred that the term "contour current" be applied only to currents flowing parallel to bathymetric contours, and other currents be termed bottom currents. For example, wind-driven surface currents may flow in a circular motion (Fig. 1) and form eddies that reach the deep-sea floor, such as the Loop Current in the Gulf of Mexico, and the Gulf Stream in the North Atlantic. Local bottom currents that move up- and downslope can be generated by tides and internal waves, especially in submarine canyons. These currents are quite capable of erosion, transportation, and redeposition of fine-to-coarse sand in the deep sea. *See* GULF STREAM; OCEAN CIRCULATION.

Bottom currents (1) generally persist for long time intervals and can develop equilibrium conditions;

(2) transport sand primarily by traction (bedload movement-sliding), rolling, and saltation; (3) are sometimes free of sediment, and are termed clear-water currents; (4) entrain and transport passive sediment particles; (5) are driven by thermohaline, wind, wave, or tidal forces; and (6) commonly flow parallel to the strike of the regional slope (**Fig. 4**) but can also flow in circular motions (gyres) unrelated to the slope. These characteristics clearly discriminate deep-sea bottom currents from turbidity currents. Bottom currents operate parallel to the strike of the slope in most deep-marine settings independently of downslope turbidity currents, debris flows, and slumps. As a result, bottom currents can constantly rework sands introduced into the basin episodically by downslope gravity processes. *See* TIDAL POWER.

Deposits of contour currents (bottom currents) have been termed contourites. However, the general term "bottom-current-reworked sands" is preferred for all types of reworking in the deep sea.

General characteristics of bottom-current-reworked sands are:

- Fine-grained sand and silt lithofacies
- Thin-bedded to laminated sand (usually less than 5 cm or 2 in.) in deep-marine mud
- Rhythmic occurrence of sand and mud layers
- Sharp (nonerosional) upper contacts and sharp-to-gradational bottom contacts
- Internal erosional surfaces
- Well-sorted sand and low depositional matrix (clean sand)
- Inverse size grading (coarsening upward) at various scales
- Horizontal lamination and low-angle cross lamination
- Cross bedding
- Lenticular bedding/starved ripples
- Current ripples and wave ripples
- Mud offshoots in ripples
- Mud-draped ripples
- Alternating traction and suspension structures
- Absence of associated thin units with normal grading
- Double mud layers (tidal)
- Sigmoidal cross bedding (tidal)
- Flaser bedding
- Occurrence of sand layers with traction structures in discrete units, but not as part of a vertical sequence of structures, such as the Bouma Sequence with basal graded division
- Lenticular-to-sheetlike geometry

Pelagic and hemipelagic settling. Pelagic and hemipelagic processes generally refer to settling of mud fractions derived from the continents and shells of microfauna down through the water column throughout the entire deep-ocean floor (Fig. 1). Hemipelagites are deposits of hemipelagic settling of deep-sea mud in which more than 25% of the fraction coarser than 5 micrometers is of terrigenous, volcanogenic, or neritic origin. Although pelagic mud and hemipelagic mud accumulate throughout the entire deep-ocean floor, they are better preserved in parts of abyssal plains (Fig. 1). Rates of sedmentation vary from millimeters to greater than 50 cm (20 in.) per 1000 years, with the highest rates on the upper continental margin.

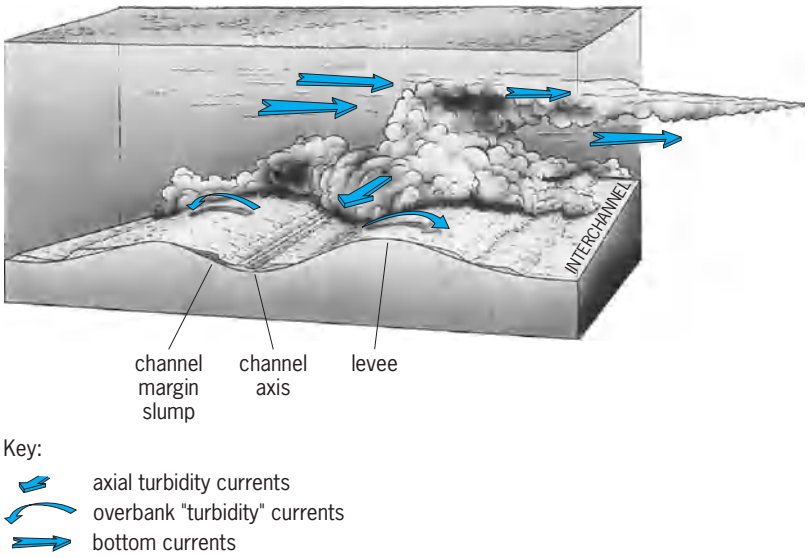
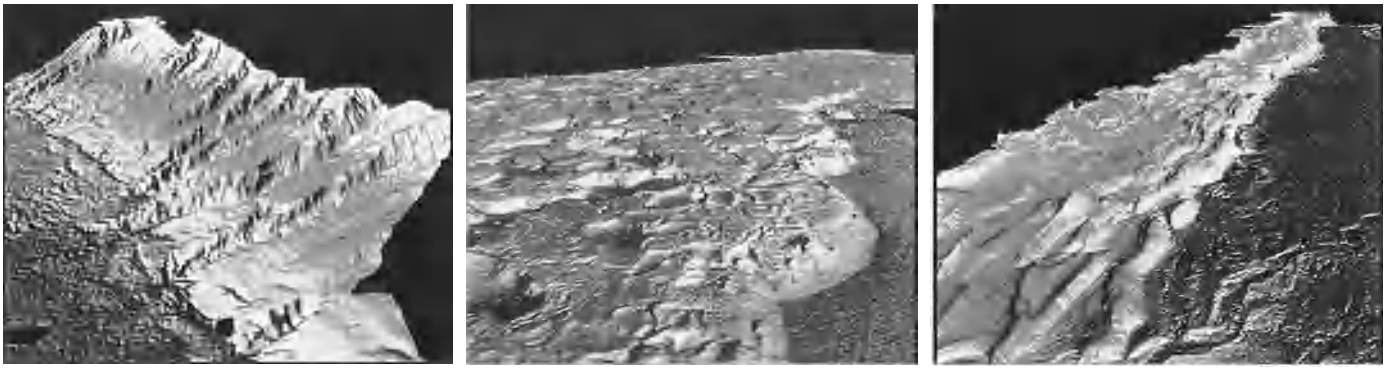


Fig. 4. Conceptual model showing interaction between channelized (axial) turbidity currents, overbank turbidity currents, and bottom currents (contour currents) in deep-sea environments. (*From G. Shanmugam, T. D. Spalding, and D. H. Rofheart, Process sedimentology and reservoir quality of deep-marine bottom-current reworked sands (sandy contourites): An example from the Gulf of Mexico, Amer. Ass. Petrol. Geol. Bull., 77:1241–1259, 1993*)

General characteristics of pelagites and hemipelagites are:

- Mud lithofacies
- Parallel lamination
- Faint normal grading
- Bioturbation
- Deep-marine body fossils and trace fossils
- Sheetlike geometry conformable to underlying sea-floor topography (drape)

Submarine slope environments. Submarine slopes are considered to be of the sea floor between the shelf-slope break and the basin floor (Fig. 1). Modern continental slopes around the world average 4°, but slopes range from less than 1° to greater than 40°. Slopes of active margins (for example, California and Oregon, about 2°) are relatively steeper than those of passive margins (for example, Louisiana, about 0.5°; **Fig. 5**). On constructive continental margins with high sediment input, gravity tectonics involving salt and shale mobility and diapirism forms intraslope basins of various sizes and shapes (for example, Gulf of Mexico; Fig. 2). Erosional features, such as canyons and gullies, characterize intraslope basins. Deposition of sand and mud occurs in intraslope basins. Slope morphology plays a major role in controlling deep-marine deposition through (1) steep versus gentle gradients, (2) presence or absence of canyons and gullies, (3) presence or absence



(a) (b) (c)

Fig. 5. Modern United States continental slope showing variability in slope values among different areas: (*a*) New Jersey-Maryland (2.5°); (*b*) Louisiana (0.5°); and (*c*) Oregon (2.0°). (*From L. F. Pratson and W. F. Haxby, What is the slope of the U.S. continental slope?, Geology, 24:3–6, 1996*)

of intraslope basins, and (4) influence of salt tectonics. *See* DIAPIR; EROSION.

Submarine canyon and gully environments. Submarine canyons and gullies are erosional features that tend to occur on the slope. Although canyons are larger than gullies, there are no standardized size criteria to distinguish between them. Submarine canyons are steep-sided valleys that incise the continental slope and shelf. They serve as major conduits for transporting sediment from land and the continental shelf to the basin floor. Modern canyons are relatively narrow, deeply incised, steeply walled, often sinuous valleys with predominantly V-shaped cross sections. Most canyons originate near the continental shelf break and generally extend to the base of the continental slope. Canyons commonly occur off the mouths of large rivers such as the Hudson and Mississippi, although many others, such as the Bering Canyon in the southern Bering Sea, have developed along structural trends. *See* BERING SEA.

Modern submarine canyons vary considerably in their dimensions. Their average length of canyons has been estimated to be about 55 km (34 mi), although the Bering Canyon, the world's longest, is nearly 1100 km (684 mi). The shortest canyons are those off the Hawaiian Islands, with average lengths of about 10 km (6.2 mi).

The water depths in which heads of canyons begin vary from a few meters (near the shoreline) to several hundred meters (shelf break and upper slope). Canyons on the Pacific margin develop at shallower depths than canyons on the Atlantic margin. For example, heads of California canyons begin at an average depth of about 35 m (115 ft), whereas canyons of the east coast of the United States begin at depths greater than 100 m (328 ft). The average depth of canyon termination has been estimated to be 2000 m (6562 ft). The average relief of canyon walls is over 900 m (559 ft). The submarine canyon with the greatest relief is the Great Bahama Canyon, where wall relief is up to 4285 m (14,058 ft).

Physical and biological processes common in submarine canyons are mass wasting, turbidity currents, bottom currents, and bioerosion. Mass wasting is a general term used for the failure, dislodgment, and downslope movement of sediment under the influence of gravity. Common examples of mass wasting are slides, slumps, and debris flows. Major slumping events can lead to formation of submarine canyons. The Mississippi Canyon in the Gulf of Mexico, for example, is believed to have been formed by retrogressive slumping during the Late Pleistocene fall in sea level. During the Holocene rise in sea level, it was partially filled. *See* HOLOCENE; MASS WASTING; PLEISTOCENE.

Francis Shepard made significant contributions to understanding the processes that operate and modify modern submarine canyons. His current-meter records from various submarine canyons around the world revealed that bottom currents (rarely exceeding 50 cm/s or 25 in./s) flow almost continuously up or down canyon axes. According to Shepard, tidal and internal wave forces are the major causes for the semidiurnal up- and downcanyon flow reversals observed in submarine canyons. Such currents can move sediment up and down canyons.

Sea-level changes and tectonic settings are important controlling factors in the development of submarine canyons. Geologic evidence suggests that downcutting of submarine canyons by sediment-gravity flows took place primarily during periods of glacially lowered sea levels. *See* GLACIOLOGY.

Submarine channel environments. According to E. Mutti and W. R. Normark, "A channel is the expression of negative relief produced by confined turbidity current flow, and represents a major, long-term pathway for sediment transport." In other words, all submarine channels are of turbidity current origin, and they all must be long-lived in order to qualify as channels. However, it is not always possible to establish whether a channel in the rock record was cut by a turbidity current or by some other process because processes that cut channels are not necessarily the same processes that fill channels. For example, many channels are filled with deposits of debris flows; however, debris flows are generally not capable of forming erosional channels. Another problem is determining whether an ancient channel acted as a pathway for a long or a short period of time. There are no standardized definitions of long and short periods

of time in deep-water sedimentation. Dimensions of submarine channels vary highly in their widths and depths.

Submarine channels can be erosional, aggradational, or both. Erosional channels commonly develop on the slope, whereas aggradational channels with levees tend to develop on basin floors having gentle gradients. Again, recognition of the type of submarine channel is not easy. Also, there are no objective criteria to distinguish erosional canyons and gullies from erosional channels. Many channels that began as an erosional feature became an aggradational feature. Overbank sediments are characteristic of the aggradational type, and are fine-grained, thin-bedded, and current-laminated. Graded mudstones are dominant in overbank deposits. Persistent overbanking can result in the development of positive relief in the form of levees.

Submarine channels on mature passive-margin fans tend to be relatively long, bifurcating, low-gradient, and largely sinuous. By contrast, channels on active-margin fans are short, steep, and low-sinuosity. The relatively fine-grained (mud-rich) character of the transported sediment associated with channels on mature passive-margin fans, such as the Amazon and Mississippi, gives rise to excellent bank stability and favors the development of a single, largely sinuous commonly meandering channel. Channel shifting in such a system is probably mainly by periodic changes in course (avulsion). However, the origin of sinuous and meandering channels in submarine setting is still poorly understood because, unlike prolonged fluvial currents that develop sinuous channel morphology on land, turbidity currents are episodic events. Therefore, it is difficult to envision how these events could develop the highly meandering channel plan forms observed in the deep sea. Some sinuous channels have been attributed to structural origin (that is, they are controlled by fault motions). *See* GEOMORPHOLOGY.

The sand-rich character and steep gradient of active-margin settings should, by contrast, favor development of a braided channel system. A long-range side-scan sonar (GLORIA) survey of the Orinoco deep-sea fan (mixed setting), which abuts against the deformation front of the Barbados Outer Ridge, also reveals a braided distributary system. *See* SONAR.

High-amplitude reflection packets. In the Amazon Fan in the equatorial Atlantic, channel bifurcation through avulsion is thought to initially lead to deposition of unchannelized sandy flows in the interchannel area (**Fig. 6***a*). Subsequent channel and levee development and progradation over these sandy deposits (Fig. 6*b*) produces a sheetlike geometry at the base of the new channel-levee system that returns high-amplitude reflection packets (HARPs) on seismic data. These sheetlike HARPs overlain by a channel-levee system (gull-wing geometry) are in many ways identical in appearance to a basin-floor fan overlain by a slope fan in a sequence-stratigraphic framework. However, there is a major difference between a basin-floor fan and a HARP. For example, a basin-floor fan is formed by progradation primarily during lowstands of sea level (allocyclic process), whereas HARPs are the results of channel bifurcation (autocyclic process). More importantly, the basin-floor fan and the slope fan are not contemporaneous, whereas a HARP and its overlying channel-levee system are essentially contemporaneous features. This again illustrates that caution must be exercised in interpreting seismic geometries in terms of processes. *See* SEISMIC STRATIGRAPHY; SEQUENCE STRATIGRAPHY.

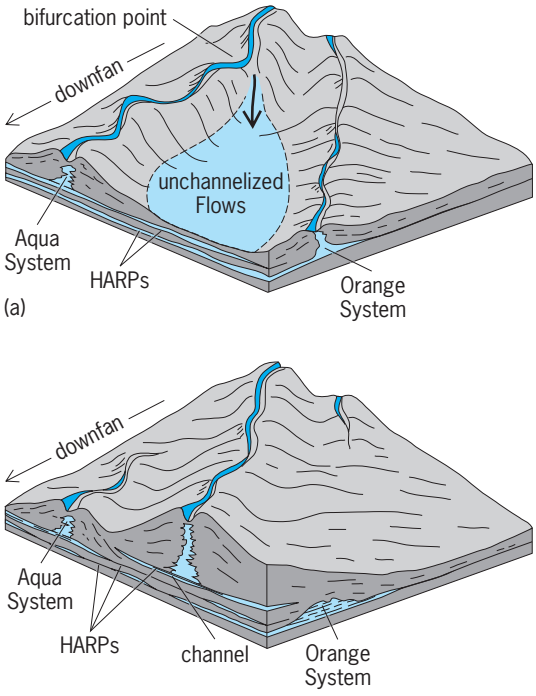


Fig. 6. Amazon Fan. (*a*) Channel bifurcation through avulsion on the deep-sea Amazon Fan results in unchannelized sandy flows by breaching their confining levee through a crevasse and spreading out initially as unchannelized flows into a lower interchannel areas. (*b*) Reestablishment of a new channel above these sandy deposits can result in a sheetlike geometry that returns high-amplitude reflections (HARPs) on seismic data. (*From R. D. Flood et al., Seismic facies and Late Quaternary growth of Amazon submarine fan, in P. Weimer and M. H. Link, eds., Seismic Facies and Sedimentary Processes of Submarine Fans and Turbidite Systems, pp. 415–433, Springer-Verlag, New York, 1991*)

Submarine-fan environments. Submarine fans represent fan-shaped or lobate deposits located in front of submarine canyons or channels. Submarine fans are considered to be formed primarily by turbidity currents coming from a point source. However, not all submarine canyons or channels develop submarine fans. Submarine-fan models, based on turbidite concepts, have been the most influential sedimentologic tools in the petroleum industry for interpreting deep-marine environments.

Sedimentologic fan models. Normark presented the first widely used sedimentologic model for modern submarine fans based on studies of small, sand-rich fans such as the San Lucas and Navy fans off California. He introduced the term "suprafan" to describe the

lobe-shaped bulge on modern fans found immediately downfan of the termination of the major feeder channel. This morphologic feature was presumably formed by rapid deposition of coarse sediment by turbidity currents at the termination of the upper-fan valley. The suprafan lobe was thought to exhibit an overall mounded, hummocky morphology in bathymetric data.

Mutti and Ricci Lucchi proposed submarine-fan models based on outcrop studies in Italy and Spain that popularized the concept of submarine fans with channels in the middle-fan setting and depositional lobes (lobate deposits) in the lower-fan setting. The general characteristics of the depositional lobes of ancient submarine fans include the following: (1) they are considered to develop at or near the mouths of submarine-fan channels analogous to distributary mouth bars in deltaic systems; (2) they show an absence of basal channeling; (3) they usually display thickening-upward depositional cycles composed of classic turbidites; (4) their common thickness range is 3–15 m (9.8–49 ft); and (5) they exhibit sheetlike geometry.

Roger Walker combined the major elements of Normark's model for modern fans with facies concepts of ancient submarine fans of Mutti and Lucchi, and advocated a general fan model with a single feeder channel in the upper-fan area and suprafan lobes in the middle/lower fan areas. Subsequently, this general fan model became influential in hydrocarbon exploration and production because of its predictive capabilities.

Sequence-stratigraphic fan models. Based partly on the suprafan-lobe model of Normark, seismic facies and geometries are used to classify deep-marine systems into basin-floor fans and slope fans in a sequence-stratigraphic framework, and in turn these models are used to predict deposits of specific depositional processes (for example, turbidity currents). However, "turbidity current" has precise meanings in terms of rheology (newtonian) and flow state (turbulence). Evidence for newtonian rheology and flow turbulence cannot be established directly from seismic-reflection profiles or wireline-log motifs; rather, these properties can be ascertained only from actual sediment facies in cores or outcrops. This is because depositional features in cores and outcrops are only centimeters in scale, and these features are too small to be resolved in seismic data. Furthermore, the interpretation of specific seismic facies and geometries (for example, sheet, mounded, continuous, hummocky) as to sediment processes may vary from one worker to the next depending on one's experience. More importantly, a single depositional facies (sandy debris flows) can generate a variety of external seismic geometries (for example, mounded, sheet, and lateral pinchout) and internal seismic reflection patterns (for example, bidirectional downlap, hummocky/chaotic, and parallel/continuous).

Abandonment of submarine-fan models. By the early 1980s, so many deep-marine sedimentary systems were studied that the simple submarine models discussed earlier were proved to be obsolete. Because the morphologic characteristics of modern suprafan lobes are either not preserved in the rock record or cannot be planimetrically mapped in outcrops, Normark abandoned his "suprafan lobe" concept altogether. Walker also recently abandoned his general fan model as "a submarine fan model of the channel-depositional lobe type, influential in its time, but now obsolete because it ignored external controls, especially sea level fluctuations."

The conceptual basin-floor fan model, characterized by mounded seismic facies, predicts sheetlike turbidite sands. However, one study provides a detailed description and interpretation of about 3600 m (12,000 ft) of core through a number of mounded seismic forms of "basin-floor fans" in the North Sea and Norwegian Sea. In these mounded features, turbidites are extremely rare (<1%). Mass-transport deposits, especially slumps, slides, and debris flows, are predominant in the core (50–100%) taken from mounded seismic facies. Recent data also suggest that some of these sands are laterally discontinuous. While features identified as basin-floor fans may occur at specific and predictable stratigraphic positions within a depositional sequence and produce characteristic seismic facies and reflection patterns on seismic data, this core study indicates that basin-floor fans do not represent specific depositional facies (for example, turbidites) and geometries (that is, sheetlike) as the model predicts. Accordingly, seismic mound models (for example, basin-floor fan) should be abandoned because they are based on a model (suprafan-lobe) that is no longer used.

Submarine basin plain environments. The term "abyssal plain" refers to a flat region of the modern deep ocean floor (**Fig. 7**), where the gradient is less than 1:1000. The geologic community, in referring to ancient examples, commonly uses the more general term "basin plain." Basin plains form in response to filling and leveling sea-floor topography by ponding of turbidity currents and by other processes. Because of their flat nature, basin plains favor sheetlike geometries for deposits of turbidity currents and other suspension settling processes (pelagic and hemipelagic). Basin plains are also potential areas for sheetlike slides and debris flows. The Enderby abyssal plain occupies a vast area covering nearly 3.7×10^6 km^2 (1.4×10^6 mi^2; Fig. 7).

Paradigm shift. Although the turbidite paradigm began in the 1950s, understanding of depositional processes in deep-marine environments is still in its infancy. No single facies model can adequately represent all deep-marine sedimentary systems. Submarine fan studies are presently in a state of flux. The days of interpreting complex deep-marine sequences as channels and lobes using fan models are over. Sedimentologic and sequence-stratigraphic lineages of fan concepts dominated by lobes show that their popularity escalated in the 1970s and 1980s but declined in the 1990s to the point of abandonment. The turbidite paradigm reached full circle in the 1990s, completing a remarkable scientific journey. Things are much improved in terms of available marine geological data, core, and outcrop studies,

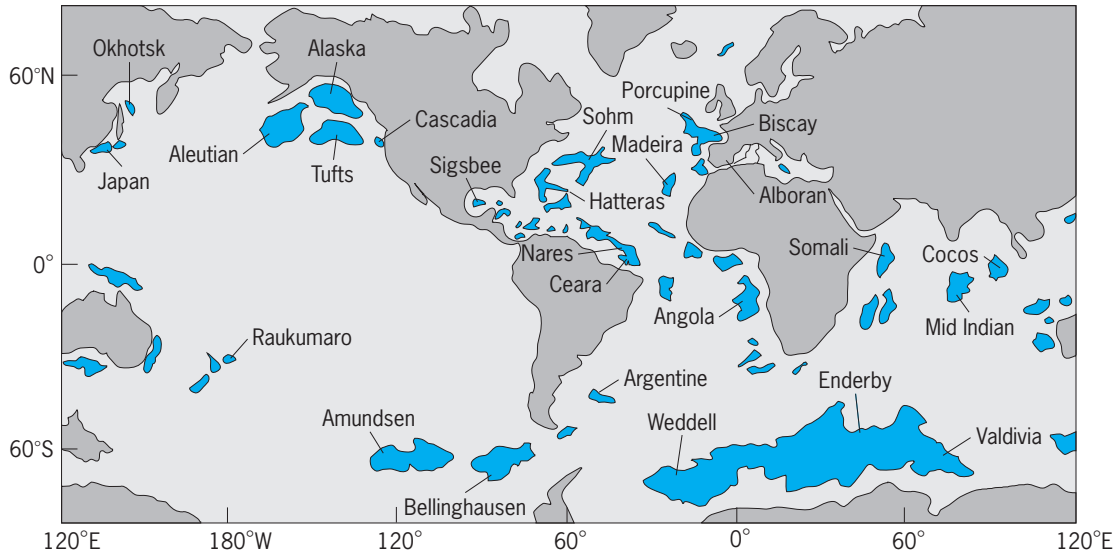


Fig. 7. Sizes of modern abyssal plains. (*From P. P. E. Weaver, J. Thomson, and P. M. Hunter, Introduction, in P. P. E. Weaver and J. Thomson, eds., Geology and Geochemistry of Abyssal Plains, Geol. Soc. London Spec. Publ., no. 31, pp. vii–xii, 1987*)

theoretical considerations, and flume experiments. Deep-marine systems are quite complex in terms of sea-floor topography, depositional processes, geometries, and stacking patterns. As a result, no single facies model can possibly explain all variations in the complex deep-sea environments.

Mass transport processes (slides, slumps, sand flows, and debris flows) have been observed in modern oceans; however, convincing direct observations of turbidity currents in modern oceans are lacking. It is ironic that there are numerous deep-marine facies models for deposits of turbidity currents that have never been observed, but there are no facies models for deposits of mass flows that are observed. This is perhaps due to the simplicity of turbidity current concepts and submarine-fan models, and the historical association between turbidites and sheet geometries. It is true that basinal turbidites are sheetlike in geometry; however, these turbidite sands are commonly thin-bedded and fine-grained and contain large amounts of mud. In comparison to turbidites, slope sands of debris flow origin are thicker-bedded and coarser-grained and contain lower amounts of mud. The current trend in the petroleum industry is to routinely apply submarine-fan models, developed for base-of-slope settings with smooth sea floors, to intraslope settings with highly irregular sea floors such as in the Gulf of Mexico. However, there is a need to develop separate models for slopes emphasizing slope processes and products. The conventional wisdom that slopes are areas of "bypassing" of sand is not always valid. Slopes are important future target areas where major petroleum reservoirs are waiting to be found. Some slopes undergo extensive gravity tectonic deformation, which leads

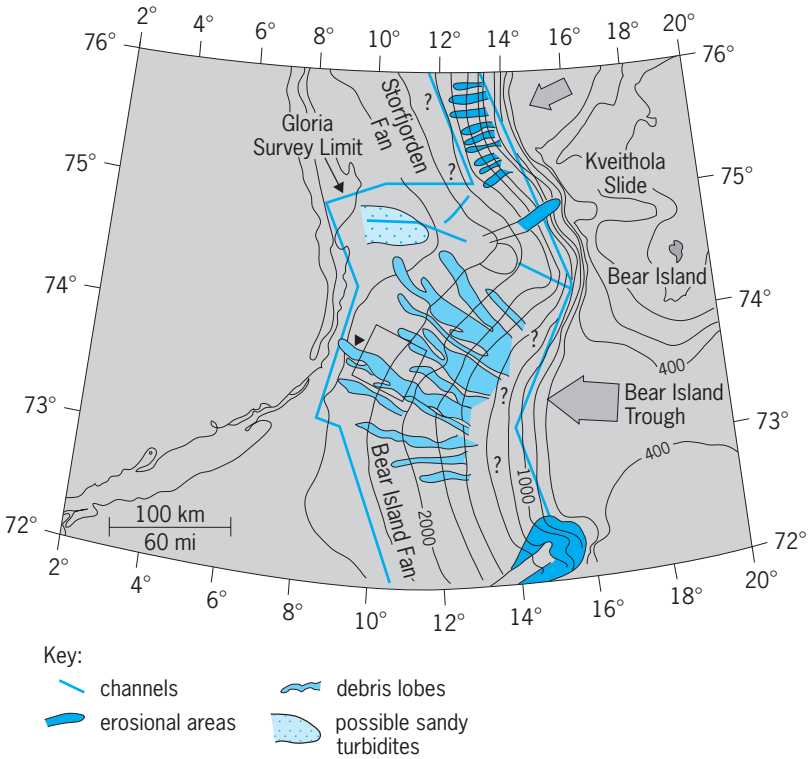


Fig. 8. GLORIA long-range side-scan sonar imagery showing fingerlike patterns of debris flows in the Bear Island Fan on the continental margin of the modern Norwegian-Barents Sea. These fans are composed of cohesive debris flows. Properties of the Bear Island Fan are: runout distance 100–200 km (62–124 mi); width 5–25 km (3–16 mi); thickness 10–50 m (33–164 ft); volume 10–30 km³ (2.4–7.2 mi³); slope angle—midslope 0.5°, lower slope 0.2°. (*From A. Elverhoi et al., On the origin and flow behavior of submarine slides on deep-sea fans along the Norwegian-Barents Sea continental margin, Geo-Mar. Lett., 17:119–125, 1997*)

to development of diapirs and intraslope basins with erosional features, such as canyons and gullies (for example, northern Gulf of Mexico, Niger Delta). Deposition of thick sand bodies can occur in such intraslope basins. Future models should take into account the great wealth of information available on modern and ancient slope processes and products such as ancient sandy slides, muddy slides, slumps, and debris flows. Debris flows can travel hundreds of kilometers on gentle gradients. On continental margins, debris flows tend to develop fingerlike patterns (**Fig. 8**). These details are seldom included in popular deep-marine models. *See* PETROLEUM ENGINEERING.

A new paradigm for deep-marine systems will emerge that will be more inclusive in terms of debris flows and bottom currents than just turbidity currents and fan models. This paradigm shift will emphasize that deep-marine systems are extremely complex and each case is unique. G. Shanmugam

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Deep-sea fauna

The deep sea may be regarded as that part of the ocean below the upper limit of the continental slopes (**Fig. 1**). Its waters fill the deep ocean basins, cover about two-thirds of the Earth's surface, have an average depth of about 12,000 ft (4000 m), and provide living space for communities of animals that are quite different from those inhabiting the land-fringing waters which overlie the continental shelves (neritic zone). *See* ECOLOGICAL COMMUNITIES.

The systematic exploration of the deep sea began with the voyage of the HMS *Challenger* (1872–1876). Since that time there have been numerous large-scale, deep-sea expeditions. In 1948 the Swedish Deep Sea Expedition in the Atlantic developed new techniques for trawling and the winch used by later expeditions. *See* MARINE BIOLOGICAL SAMPLING.

The deep-sea fauna consists of pelagic animals (swimming and floating forms between the surface and deep-sea floor) and below these the benthos, or bottom dwellers, which live on or near the ocean bottom. Pelagic animals can be divided into the usually smaller forms that tend to drift with the currents (zooplankton) and the larger and more active nekton, such as squids, fishes, and cetaceans. Pelagic, deep-sea animals are frequently termed bathypelagic in contrast to the epipelagic organisms of the surface waters (Fig. 1).

Bathypelagic fauna. All animal life in the sea, pelagic and benthic, depends on the growth of microscopic plants (phytoplankton). From the surface down to a maximum depth of about 300 ft (100 m) there is sufficient light for photosynthesis and vigorous phytoplanktonic growth. This layer is known as the photic zone. In the deep sea, plants can exist only as saprophytes. The productivity of the plants, however, is reflected down to the deepest parts of the sea through complex food chains. These consist of zooplankton that graze on phytoplankton, carnivorous species that feed on zooplankton, and large predators that eat the other animals. The typical bathypelagic animals (**Fig. 2**) begin to appear below depths of about 600 ft (200 m). *See* FOOD WEB; PHYTOPLANKTON; ZOOPLANKTON.

Zooplankton. The planktonic or drifting forms of animal life in the ocean include the Protozoa, larval stages of deep-sea fishes, and even larger organisms with limited powers of movement.

1. *Protozoa*. Included in this group are various families of Foraminiferida and Radiolaria, the skeletons of which form an important part of the deep-sea sediments. *See* PROTOZOA.

2. *Coelenterata*. Scyphomedusae such as *Atolla* and *Periphylla* are not uncommon. Other jellyfishes include various Trachymedusae (*Crossota* and *Colobonema*) and Narcomedusae. Siphonophorida, particularly the diphyids, are found down to depths of at least 9000 ft (3000 m) but are more common in the upper several hundred feet. *See* CNIDARIA.

3. *Nemertea*. This group of worms has bathypelagic species belonging to some 10 families.

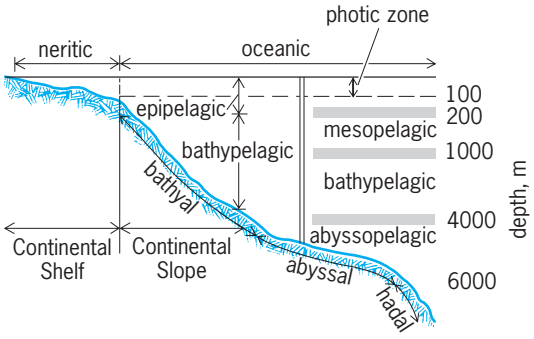

Fig. 1. Classification of marine environments. Right side of diagram illustrates the proposal to divide the bathypelagic zone into mesopelagic, bathypelagic, and abyssopelagic zones. Division of benthic region into bathyal, abyssal, and hadal zones also is shown. 1 m = 3.3 ft.

4. *Crustacea*. In numbers of species and individuals, the small Copepoda are the dominant group of crustaceans in the ocean. There are numerous bathypelagic species. Certain of the Ostracoda (*Gigantocypris*) are purely bathypelagic, as are some of the Amphipoda (the gammarid genera *Cyphocaris* and *Hyperiopsis* and most species of the hyperiid families Scinidae and Lanceolidae). *See* AMPHIPODA; COPEPODA; OSTRACODA.

The larger and more active pelagic crustaceans (Euphausiacea, various Mysidacea, and prawns) are usually classed as plankton but might well be called "micronekton," a group intermediate between thrusting nekton and feebler-swimming plankton. Of the euphausiid shrimps, *Benthenphausia* and various species of *Thysanopoda*, *Nematoscelis*, and *Stylocheiron* have centers of abundance in the bathypelagic zone. Deep-water genera of mysids include *Gnathophausia*, *Lophogaster*, and *Eucopia*, while the prawn families Hoplophoridae and Sergestidae have numerous bathypelagic representatives. *See* CRUSTACEA.

5. *Chaetognatha*. Certain species of arrowworms, such as *Eukrohnia fowleri* and *Sagitta macrocephals*, are predominantly bathypelagic. *See* CHAETOGNATHA.

6. *Echinodermata: Holothuroidea*. The genera *Pelagothauria*, *Enypiastes*, and *Galatheathuria* are bathypelagic, the first two being medusalike. *See* ECHINODERMATA.

7. *Protochordata: Thaliacea*. While they are more abundant in the surface layers, the salps, doliolids, and pyrosomes have been fished down to 9000 ft (3000 m).

Deep-sea nekton. This group consists largely of squids, octopods, and fishes. The sperm whale also enters the deep sea, where it finds some of the squid upon which it feeds.

1. *Mollusca: Cephalopoda*. Several families of squids form an important part of the deep-sea nekton together with a few octopods, such as *Cirrothauma*, *Amphitretus*, *Vitreledonella*, and *Vampyroteuthis*. *See* CEPHALOPODA.

2. *Fishes*. Apart from a few squaloid sharks, the bathypelagic fish fauna consists of teleosts. The most diverse groups are the stomiatoids (Elupeiformes), with about 300 species; Myctophidae (lantern fishes; Salmoniformes), about 250 species; and the ceratioid angler fishes (Lophiiformes), about 90 species. The few species forming the orders Anguilliformes (gulper eels) and Cetomimiformes (whale fishes) are entirely bathypelagic, as are certain of the eels (Cyemidae, Nemichthyidae) and Beryciformes (for example, Melamphaidae). *See* TELEOSTEI.

Distribution. The bathypelagic fauna is most diverse in the tropical and temperate parts of the ocean. Numerous species are found in all three temperature zones, but many appear to have a more limited distribution.

Each species also has a definite vertical occurrence. Findings suggest that there are three main vertical zones, each with a characteristic community. Here the term bathypelagic is used for the fauna between about 3000 and 6000 ft (1000 and 2000 m), that above (between 600 and 3000 ft or 200 and 1000 m) being called mesopelagic and that below 6000 ft (2000 m) abyssopelagic (Fig. 1). The typical forms of the mesopelagic fauna (stomiatoids and lantern fishes) live in the twilight zone of the deep sea (between the 68 and 50°F or 20 and 10°C isotherms), while the bathypelagic species (ceratioid angler fishes and *Vampyroteuthis*) occur in the dark, cooler parts below the 50°F (10°C) isotherm.

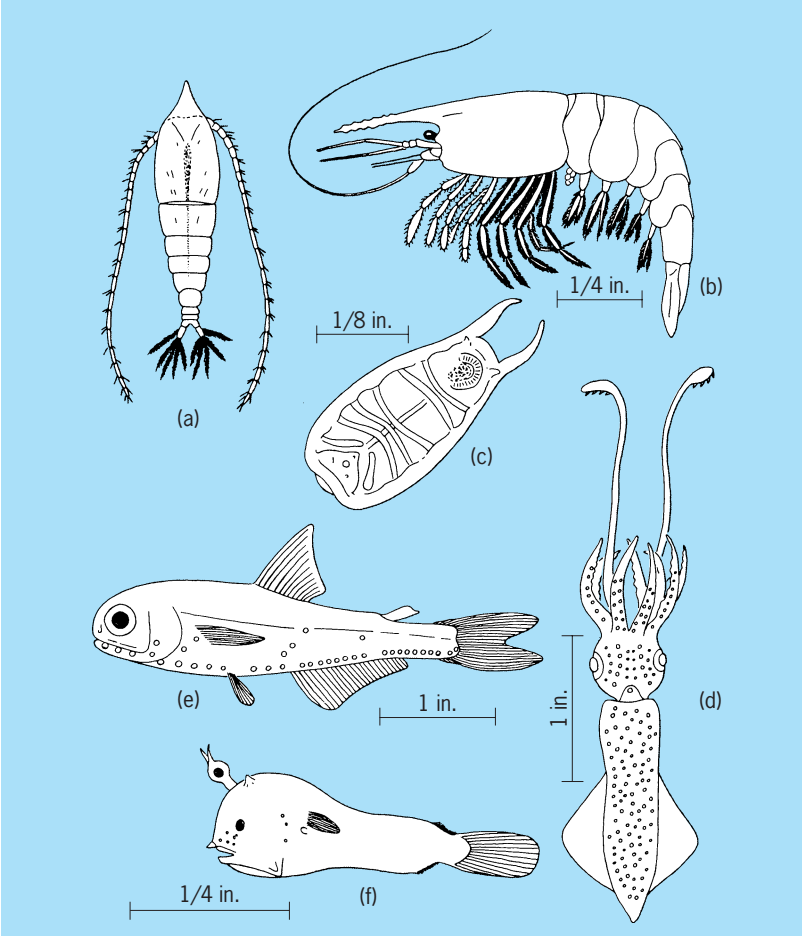

Fig. 2. Pelagic animals of the deep sea. (*a*) Copepod (*Haloptilus acutifrons*). (*b*) Prawn (*Acanthephyra multispina*). (*c*) Salp (*Salpa*, or *Thalia*, *democratica*). (*d*) Squid (*Abraliopsis morisii*). (*e*) Lantern fish (*Myctophym punctatum*). (*f*) Anglerfish (*Lophodolus acanthognathus*).

Lastly, numerous species of mesopelagic animals, such as euphausiids, prawns, squids, and fishes (particularly lantern fishes), undertake extensive diurnal, vertical migrations, moving upward into the productive surface layers to feed at night. Toward sunrise they begin to descend to their daytime levels. *See* SCATTERING LAYER.

Bioluminescence. Perhaps the most conspicuous feature of pelagic deep-sea life is the widespread occurrence of luminescent species bearing definite light organs (photophores). Many of the squids and fishes have definite patterns of such lights, as do some of the larger crustaceans (hoplophorid and sergestid prawns and euphausiids). Investigations have shown that flashes from luminescent organisms could be

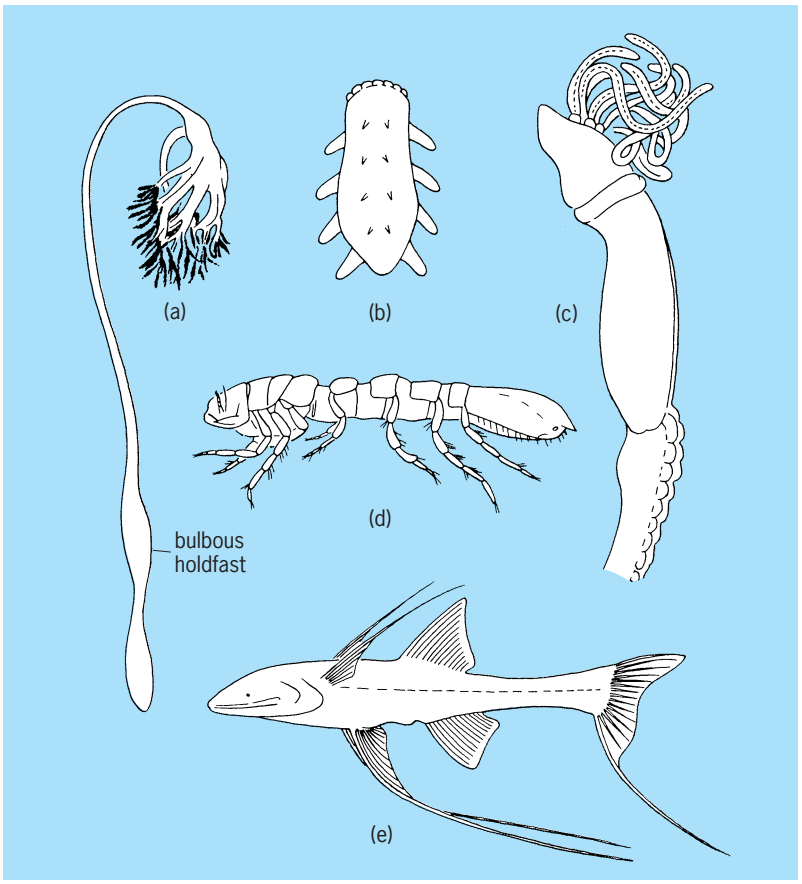

Fig. 3. Benthic animals of the deep sea. (*a*) Sea-pen (*Umbellula*) showing bulbous holdfast. (*b*) Sea cucumber (*Elpidia glacialis*). (*c*) Head and forepart of the trunk of a pogonophoran (*Birsteinia witjasi*). (*d*) Isopod crustacean (*Macrostylus hadalis*). (*e*) Bathypteroid fish (*Benthosaurus*).

detected down to depths of 12,300 ft (3750 m). *See* BIOLUMINESCENCE; PHOTOPHORE GLAND.

Benthic fauna. There are two main ecological groups of bottom-living animals (**Fig. 3**) in the ocean: organisms that attach to the bottom and those that freely move over the bottom.

Attached benthic organisms. This group consists of species that attach themselves to the sediments, rocks, or to other organisms. The more typical forms include the following: Hexactinellida (Porifera) (glass sponges), with about 375 species; certain hydroids, gorgonians, pennatulids (sea fans), antipatharians (black corals), actiniarians (sea anemones), and madrepore corals (*Lophohelia* and *Amphihelia*); Cirripedia (barnacles), such as *Scalpellum* and *Verruca* spp.; numerous species of stalked crinoids living in the deep sea together with a number of unstalked forms; and Pogonophora (beardbearers) and certain ascidians (*Culeolus* sp.).

Benthic crawlers and swimmers. This group comprises the freely moving animals, those that swim or crawl over the bottom or burrow into the sediments, the upper layer of which has a rich bacterial flora. (1) Annelida: a few species of Polychaeta (bristle worms). (2) Gephyrea: certain species of echiuroid and sipunculoid worms. (3) Crustacea. In numbers of species and individuals the most important group of benthic deep-sea crustaceans is the Peracarida, represented by various species of cumaceans (*Bathycuma*, *Macrocylindrus*), isopods (*Ischnomesus* and *Eurycope*), amphipods, and tanaids (*Apseudes*, *Neotanais*). Of the Eucarida, the most prominent groups are the penaeid prawns and the Eryonidae. There are also a number of crabs (*Platymaia*, *Geryon*, *Ethusa*, and *Scyramathia*) and hermit crabs (numerous species of Axiidae). (4) Pycnogonida (sea spiders): numerous species of the families Colossendeidae and Nymphonidae. (5) Mollusca. Certain of the Octopodidae (octopuses) and the cirromorph octopods live on the deep-sea floor, as do various gastropods, scaphopods, and lamellibranchs. (6) Echinodermata. These form an important part of the benthic fauna, particularly the sea cucumbers (Holothuroidea) of the orders Elasipodida and Molpadida. Among the sea urchins (Echinoidea), the order Cidaroida and the suborder Meridosternata mainly consist of deep-sea species. There are also various brittle stars (Ophiuroidea) and starfishes (Asteroidea). (7) Fishes. The species of one group of cartilaginous fishes (Holocephali) live over the continental slope. The main groups of benthic, deep-sea teleosts are the Bathypteroidae (Salmoniformes); Halosauridae and Notacanthidae (Notacanthiformes); Macrouridae (rattails) and Morinae (deep-sea cods) (Gadiformes); and Brotulidae, Liparidae, and Zoarcidae (Perciformes). *See* ECHIURIDA; POLYCHAETA; PYCNOGONIDA; SIPUNCULA.

Distribution. The benthic fauna is most diverse in the temperate and tropical ocean, although the arctic and antarctic areas have their characteristic species. As in the pelagic fauna, certain species occur in all three oceanic zones, while others appear to have a more restricted occurrence.

While a number of species—particularly among the polychaete worms, gastropod mollusks, and the brittle stars (Ophiuroidea)—range from littoral to abyssal regions, most forms tend to live within smaller ranges of depth. Data suggest that there are typical communities of animals over the continental slopes (Fig. 1) extending down to about 9000 ft (3000 m; bathyal zone); others occur below this in the abyssal zone. Danish and Russian exploration also suggested that the deep-sea trenches (with depths over 21,000 ft or 7000 m) form another ecological zone (hadal zone) having certain characteristic species—those capable of living under pressures of 700–1000 atm (70–100 megapascals; barophilic species). This work also showed that life could exist at the very bottom of the ocean (down to depths of more than 30,000 ft or 10,000 m) and that species of certain groups, such as sea anemones, echiuroid and polychaete worms, bivalves, isopod and amphipod crustaceans, sea cucumbers, and Pogonophora, occurred at depths beyond 27,000 ft (9000 m).

Lastly, there is a decrease in the numbers of species and individuals with depth. Russian biologists found that at depths of 24,000–30,000 ft (8000–10,000 m)

the weight of animals per square meter of sea floor was about one-fifth to one-fifteenth the weight at depths of 3000–12,000 ft (1000–4000 m). As the deep-sea benthic fauna is dependent on organic matter originating in the upper, plant-bearing waters and as the amount reaching the bottom must decrease with depth, the above findings are comprehensible. It is also interesting that there are very few carnivorous animals, such as crabs, brittle stars, and starfishes, below a depth of 21,000 ft (7000 m). It is the particle-catchers, such as the Pogonophora, and ooze-eaters, such as sea cucumbers and echiuroid worms, that make up most of the hadal fauna. *See* MARINE ECOLOGY; SEAWATER FERTILITY.

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Deep-sea trench

A long, narrow, characteristically very deep and asymmetrical depression of the sea floor, with relatively steep sides. Oceanic trenches characterize active margins at the ocean-basin–continent or ocean-basin–island-arc boundaries. They contain the greatest oceanic depths and are associated with the most active volcanism, largest negative gravity anomalies, most frequent shallow seismicity, and almost all of the intermediate and deep-focus earthquake activity. As the surface expression of the widely accepted process of subduction by which oceanic crustal material is returned to the upper mantle, they are key elements in current models of plate tectonic evolution on Earth and possibly on Venus.

Occurrence. Deep-sea trenches are the signature relief form of the Pacific; in a counterclockwise direction, they occur from southern Chile to just northeast of North Island, New Zealand (**Fig. 1**). A secondary or outer branch trends southward from near Tokyo Bay in a festoon of arcs to south of Palau. The principal gaps in the circum-Pacific chain are from Baja California to southeastern Alaska, and off the northern coast of New Guinea. From eastern New Guinea to southern Vanuatu the trenches lie southwest, or "inside," the island chains; otherwise their characteristics are like those facing the Pacific. The Indian Ocean contains only the very long contorted Sunda Trench that appears near the northwestern end of Sumatra and extends southeast and east past Timor, to curve north and west near Aru and end adjacent to Buru. In the Atlantic, the Puerto Rico–Antillean trench system extends outside the island arc from eastern Hispaniola around to Trinidad; but south of 14°N, off Barbados, the trench is filled with sediment. In the far South Atlantic a typical island-arc–trench complex extends from near South Georgia through the South Sandwich Archipelago.

Mid-to-late-nineteenth century exploration indicated that the deepest areas of the oceans lay near continents or island arcs, in elongated depressions. Understanding of the trenches' fundamental significance to global tectonics and basin evolution has come only during the twentieth century, with the development and application of precise or indicative techniques of measurement and observation. Examples include pinpointing from land observatories the distribution and focal mechanism of deep earthquakes; laborious shipboard collections to relate geophysical data on crustal structure, gravity, magnetic patterns, and precise topography; and geological-geochemical data from igneous rocks and sediment samples obtained by dredging, coring, or drilling.

Geophysical exploration. A series of pioneering gravity observations with pendulum instruments on Dutch submarines during the 1920s and 1930s established that the East Indian trenches, and several others, were characterized by a belt of negative gravity anomalies of 150–200+ milligals, that is, values 150–200 parts per million less than normal, interpretable as deficiency of mass near and at their axes. Investigators attributed the deficiency in gravity to a symmetrical downbuckle of lighter crust into the mantle, a result of compression at unstable boundaries that might have produced the trenches. This postulated root, called a tectogene, was produced in physical model studies and was widely accepted until the development of shipboard seismic refraction techniques and more precise station seismology foci measurements after World War II.

Seismologists in California and Japan independently noted the association of great oceanic deeps such as Tonga Trench, the Peru-Chile Trench, and the Western Pacific trenches with sloping zones of shallow–intermediate–deep-focus earthquakes, and attributed a fault origin to oceanic deeps, postulating reverse faulting with the continental or island-arc side as the overthrusting member. In some areas, for example, within the "interior" Melanesian complex, the directions were reversed, but relative displacement was by thrust faulting. This gross process, as a specific mechanism for this phenomenon, had considerable appeal.

Other seismologists, interpreting the reduced gravity differently, postulated a tensional origin for trenches, modeling a thinning of the crust there. With the application of two-ship seismic refraction methods, it was established that oceanic crust is thin, that crust under island arcs is thicker, and its layers display different sound transmission velocities, indicating different composition. Shipboard studies in the Middle America, Tonga, Cedros, Aleutian, Peru-Chile, and Sunda trenches established that the characteristic oceanic crustal layer [that is, 6.8–7.0 km/s (4.1–4.2 mi/s) compressional wave velocity] does not end or thin under the trench; rather, it may thicken slightly but does deepen steeply as it passes beneath the island arc or continental slope by the process that has come to be known as subduction. The shoreward flank of the trench consists of complexes of materials of intermediate velocity, such as

Fig. 1. Deep-sea trenches in the Pacific-Melanesian-Indonesian hemisphere. Broken line indicates filled trench.

volcanics or contorted sediments. Measurements of heat flow through the sea floor on the outer flanks and on the trench floors were lower than normal for deep-sea areas, suggesting progressive cooling toward the trench, or possibly blanketing by sediments there. *See* EARTH, GRAVITY FIELD OF; EARTH, HEAT FLOW IN; EARTH CRUST; FAULT AND FAULT STRUCTURES; OCEANIC ISLANDS; SEISMOLOGY; SUBDUCTION ZONES.

Sea-floor topography and surficial processes. Topographic studies established that, in cross section, trenches are asymmetrical, with the nearshore flank the steeper, often 4–8° on the shallower foreslope becoming 10–16° on the lower slopes, and locally with precipices of 45° or more. Benches or small flat sediment ponds behind pinnacles are common, providing evidence of talus movement or mass wasting of sediment downward, probably triggered by earthquakes. Bottom photography, specifically of the Tonga Trench, shows evidence of slumps, ripple marks on sediment surfaces at great depths, and angular boulders or slickensided outcrops on the deeper reaches of the shoreward flank. The offshore flanks of trenches almost always are gentler in slope, still steepening downward, often with evidence of fracturing under tension, specifically grabens by normal faulting. Trenchward tilting of the beveled summits of at least two flat-topped seamounts on the outer slopes of the Tonga and Aleutian trenches provides direct evidence of depression and flexing on the outer slope. A low crustal swell outside the trench, required by some models, frequently is recognized. Bathymetric exploration of the trenches has been sufficient to establish a definitive ranking of their maximum depths (see **table**). *See* GRABEN; MARINE GEOLOGY; MARINE SEDIMENTS; MASS WASTING; SEAMOUNT AND GUYOT.

Crustal composition: evidence from dredging. Only with the development of ultrastrong dredging cable and stroboscopic-timed equipment-positioning devices has it been possible to collect well-located specimens of the consolidated sediment and

Trenches ranked by bathymetric depth

Trench	Maximum depth, m*
Challenger Deep, Mariana Trench, western Pacific	10,920±10
Tonga, southwestern Pacific	10,800±5
Philippine, western Pacific	10,057±5
Kermadec, southwestern Pacific	10,047±50
Izu-Ogasawara, western Pacific	9,780±10
Kuril, northwestern Pacific	9,550±5
North New Hebrides, Melanesia	9,175±10
New Britain, Melanesia	8,940±10
Yap, western Pacific	8,650±10
Puerto Rico, western Atlantic	8,605±5
South Sandwich, southern Atlantic	8,325±10
South Solomon, Melanesia	8,322±10
Peru-Chile, southeastern Pacific	8,170±10
Palau, western Pacific	8,055±10
Aleutian, northern Pacific	7,679±?
Ryuku (Nanseishoto), western Pacific	7,460±10
Sunda, northeastern Indian Ocean	7,125±10
Middle America, eastern Pacific	6,662±10

*1 m = 3.3 ft.

volcanic-plutonic igneous rocks exposed on the deep flanks of the great Pacific trenches. Meanwhile high-energy seismic reflection systems with complicated arrays and computerized data-processing methods have revealed the structural complexities—fault traces, contorted layer interfaces, rock slivers, shingling, sediment packets—that characterize other trenches' onshore flanks. There, accretion of upper crustal material from offshore by scraping, or by slumping or mass wasting of upper slope sediments, are dominant processes; evidence comprises the filled or filling trenches in the southern Caribbean Antilles, off southern Chile, off Central America, or even exposed in the islands off Sumatra by uplift.

The three deepest trenches—Mariana, Tonga, and Philippine—all display exposures or in-place outcrops of igneous rock on the deeper 5000 m (16,500 ft) of their nearshore flanks. In the Philippine Trench at its deepest off Mindanao there are volcanic rocks—basalts, basaltic andesite, and andesite transitional from arc-tholeiitic to arc-calcakaline series—as flows or dikes. These volcanics appear characteristic of the Samar–East Mindanao arc massif; there is no evidence of accretion of material from the subducted Philippine Sea oceanic crust, or even a deformed volcanoclastic apron. The nearshore flank of the Mariana Trench's Challenger Deep, the oceans' greatest abyss, also shows only rocks of island-arc affinity throughout: both andesitic and basaltic andesite volcanics, cherts, serpentinites, diabase dikes, and at lower middepths sparse plutonic granodiorites and gabbros. Again there is no evidence of accretion; apparently subduction erosion and nonaccretion prevail.

Very clear geological field evidence for a crudely layered trench-flank–island-arc foundation, for the oceanic crust being subducted, and for possible subduction erosion and nonaccretion has come from dredging in the 9500–10,000+ m deep (31,000–33,000+ ft) segment of Tonga Trench. Findings on the deepest 5000 m (16,500 ft) of both flanks are bolstered by open-ocean drilling on the upper foreslope. Dredging of 150–600-m (500–2000-ft) traverses on the nearshore flank at depths of 5200–9700 m (17,000–32,000 ft) suggest a crude layering: (1) at 5000–7000 m (16,500–23,000 ft), silicic volcanoclastics, pumice, and a siltstone carapace intruded by diabase; (2) at 7000–8500 m (23,000–28,000 ft), altered basalts overlying gabbros; (3) at 8500–9200 m (28,000–30,000 ft), coarse-grained exceedingly fresh (that is, unaltered) harzburgite and dunite; (4) locally, vesicular intermediate volcanic rocks (seamount fragment?); and (5) at 9200–10,000 m (30,000–33,000 ft), fresh to altered oceanic tholeiitic basalt (pillow fragments) from the Pacific plate crust. Two hauls on faulted offshore slopes (graben) at 8800–9000 m (29,000–30,000 ft) yielded oceanic basalt and minor diabase. An interpretive section based on these dredged rocks, seismic reflection and multibeam surveys, and seismic refraction data offshore shows oceanic crust passing beneath the peridotitic bases of the Tonga arc massif (**Fig. 2**). The precipitous lowermost flanks on both sides may represent a graben in the downflexed oceanic crust; its narrow bottom is flattened by turbidites and siliceous ooze. *See* PETROLOGY; PLUTON.

Evolution: evidence from drilling. Drilling into the sedimented forearcs (upper slopes seaward of the island chains) of several Pacific trenches elucidated characteristics of volcanism in the initiation–early subduction phases of island-arc–trench systems. A key example is the Izu-Bonin-Mariana forearc; a drilled hole in the central Tonga arc near Eua yields similar results but differs in detail. In both, the earliest volcanism, in mid to late Eocene time, constructed a wide platform of depleted (tholeiitic) arc extrusives that erupted in an extensional environment. This volumetrically great extrusion was built upon, or replaced, preexisting oceanic crust, some evidence of which remains in the Tonga forearc. Locally the forearc was subjected to rapid subsidence and subduction erosion, as evidenced by normal faulting and extension in an overall convergent system. Old lavas are found nearer the trench. With establishment of a mature subduction system (after 10,000,000+ years), normal arc volcanics (andesites and dacites) were erupted from a series of points; these evolving magmas incorporated sedimentary debris and volatile components obtained from the subducted oceanic slab. Such is the volcanism that builds the island chains. Serpentinite diapirs, produced by hydration of igneous rocks, occur at moderate depths in the Izu-Bonin forearc.Volcanic dikes of mid to late Tertiary age from island-arc magma sources bake sediments in both drilled forearcs; other sediments are hydrothermally mineralized by percolating fluids. This model, in which subduction-assimilation-magma evolution and subduction erosion are major elements, may

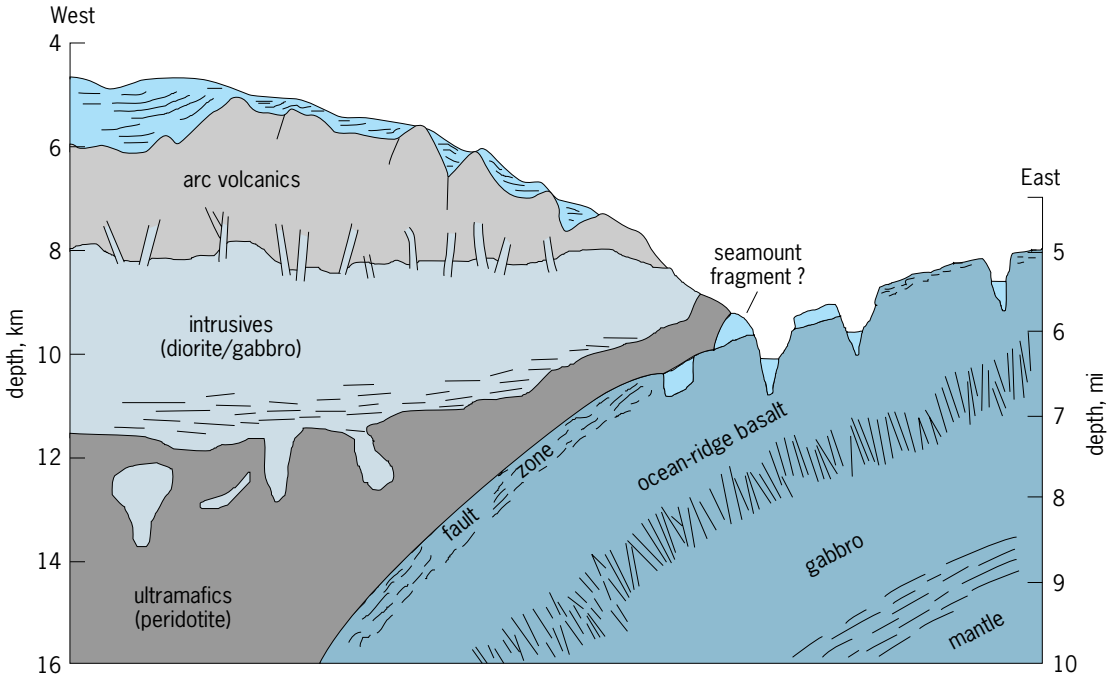

Fig. 2. Interpretive section across Tonga Trench near 20°S, based on dredged rock distribution, seismic refraction and reflection surveys, and seabeam data; vertical exaggeration 4×. (*After S. H. Bloomer and R. L. Fisher, Petrology and geochemistry of igneous rocks from the Tonga Trench: A non-accreting plate boundary, J. Geol., 95:469–495, 1987*)

characterize the sediment-starved, steady-state megatrenches of the western Pacific. Drilling and multichannel reflection profiles in regions of abundant sediment supply, for example, off Barbados or Washington State, have revealed the complicated interactions of sediment-volcanic mélanges and fluid flow in accretionary prisms, "transient" trench locales where borders advance and islands have risen at the expense of the ocean. *See* HYDROTHERMAL VENT; PLATE TECTONICS; VOLCANOLOGY.

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Deer

The name for a significant family (Cervidae) of 41 species of antler-bearing, ruminant, even-toed ungulates (Order Artiodactyla). These mammals vary in size from small and delicate to large and massive, and range from the dimensions of a rabbit to those of a large horse. Members of this family include the musk deer, Chinese water deer, barking deer, fallow deer, axis deer, sambar deer, sika, red deer, roe deer, American wapiti (elk), white-tailed deer (**Fig. 1**), mule deer, black-tailed deer, marsh deer, moose, and caribou (reindeer). The smallest living deer species is the Northern Pudu (*Pudu mephistopheles*) of Peru, Ecuador, and Colombia with a shoulder height of 32–36 cm (13–14 in.). The largest living species is the Alaskan moose (*Alces alces*) which stands 140–235 cm (4.6–7.8 ft) at the shoulder, weighs nearly a ton, and has palmate antlers that can span 2 m (6.5 ft). In former times, deer were even larger: the now extinct great Irish elk (*Megaceros giganteus*) had antlers that stretched 3.3 m (11 ft) or more from tip to tip.

Distribution and habitat. Members of the family Cervidae are spread throughout the Northern Hemisphere, including Asia, Europe, and North America. In the Southern Hemisphere, however, their range is limited almost entirely to South America. With the exception of certain kinds of red deer that are found in the Barbary Coast area, this family is absent from Africa, and no native species exist in Madagascar or Australia (see **table**). Deer may inhabit lowland swamps, grassy plains, and sparsely covered brush country as well as dense woodland. They are found in mountainous terrain and even in the tundra. They may feed on grasses, herbs, leaves, buds, shoots, fruit, twigs, bark, fungi, mosses, and lichens. Most species are more or less solitary, but can occur in small groups or herds. Most species are not territorial. Some undertake regular, annual migrations.

Morphology. Deer have short tails, large ears, and a muzzle that ranges from being naked to almost entirely covered with hair. The pelage, which varies

Fig. 1. Male white-tailed deer (*Odocoileus virginianus*).

from being short and smooth to dense and long, ranges in color from light gray to brown to black. Skin glands, such as preorbital, interdigital, tarsal, and frontal, are common. The feet have four toes: the third and fourth digits are well-developed, whereas the second and fifth are small and do not touch the ground. The dental formula is I 0/3, C 1/1, PM 3/3, M 3/3 × 2 for a total of 34 teeth. Deer have no central incisors in their upper jaw, their place being taken by a pad of hard gum tissue. Enlarged canines are present in musk deer (*Moshus moschiferus*), Chinese water deer (*Hydropotes inermis*), muntjac deer (*Muntiacus* sp.), and tufted deer (*Elaphodus cephalophus*). The body length in deer ranges 80–310 cm (2.6–10.3 ft) with the tail length ranging 1–50 cm (0.4–20 in.) and a weight ranging from 7 to about 800 kg (15 to about 1760 lb).

Antlers. Sexual dimorphism is evident in most species of deer with the most characteristic feature being the presence of antlers on the males of most species. Antlers are one of the most remarkable features in the entire animal kingdom. Generally, they are found only on the male. Exceptions include the musk deer (**Fig. 2**) and Chinese water deer, which lack antlers in both sexes, and the caribou in which both males and females possess antlers. Antlers are solid bony growths that develop from permanent bases on the frontal bones of the skull. The deer sheds them annually, usually in midwinter or early spring; it is controlled by male sex hormones. Several weeks after the old antlers are dropped, a round fur-like ball (burr) begins to rise from each base. These small growths rapidly swell and expand into the curving and branching structures that will soon grace the head. They can grow up to 20 mm (0.8 in.) a day. During this stage, the antlers are soft and spongy and are covered with a highly vascular skin with a coat of fine hair (velvet). Growth continues until the supply of blood and nourishment slows down. Finally, it stops altogether as the blood vessels shrink and cease to function. The antlers now harden to bonelike consistency, while the overlying "velvet" dries up and peels off. The animals burnish and polish their antlers by rubbing them against tree trunks and branches—a process probably prompted by the itching sensation caused by the drying up of the living tisssues. This periodic loss and replacement of antlers every year is amazing, with even the enormous antlers of the moose, which may weigh up to 27 kg (60 lb), falling off and reappearing every year. While antlers vary quite a bit in size and shape, those of each species of deer have their typical pattern. They are used as formidable weapons against other kinds of animals and as weapons against rival males in battles for possession of females.

Rumination. Deer are ruminants. When a deer browses, fresh food is partially masticated and only coarsely broken down before being swallowed. It passes into the first (rumen) and second (reticulum) portions of the four-compartment stomach. Here the food is partially predigested by bacteria. Later, the

Representative species of Eurasian and South American deer and their geographical distributions

Species	Distribution
Musk deer *Moschus moschiferus*	Asia
Chinese water deer *Hydropotes inermis*	Eastern Asia
Montjac, or Barking deer *Muntiacus* sp. *Megamuntiacus vuquangensis*	Southern Asia
Tufted deer *Elaphodus cephalophus*	Southern Asia
Axis deer *Axis* sp.	Southern Asia, Phillipines
Fallow deer *Dama dama*	Southern Europe, Southeast Asia
Pere David's deer *Elaphurus davidianus*	China
Sambar *Cervus unicolor*	Southern Asia, Borneo, Sumatra
Barasingha *Cervus duvauceli*	Southern Asia
Red deer *Cervus elaphus*	Europe
Roe deer *Capreolus capreolus*	Europe, Asia
Pampas deer *Ozotoceros bezoarticus*	South America
Marsh deer *Blastocerus dichotomus*	South America
Guemal *Hippocamelus* sp.	South America
Pudu *Pudu* sp.	South America
Brocket deer *Mazama* sp.	South and Central America

Fig. 2. Adult male Siberian musk deer (*Moschus moschiferus moschiferus*). (© *Copyright © by Brent Huffman, 2001*)

animal can contract the rumen and reticulum and return some of the food—now called the cud—to its mouth where it is thoroughly chewed and swallowed again. This time the cud bypasses the rumen and reticulum and passes through the other two compartments (omasum and abomasum) before entering the small intestine.

Reproduction. Gestation ranges from 180 days in the muntjac to 262 days in the wapiti and usually results in the birth of one to two young annually. Muntjacs can mate twice in the same year. Weaning occurs between 2 and 12 months with sexual maturity being reached between 6 months (muntjacs) and 3 years (red deer, wapiti) of age. Life spans range from about 10 to 30 years.

North American Deer. Members of the family Cervidae inhabiting North America include the white-tailed deer (*Odocoileus virginianus*; Fig. 1), mule or black-tailed deer (*O. hemionus*), wapiti or elk (*Cervus elaphus*), moose (*Alces alces*), and caribou or reindeer (*Rangifer tarandus*). The white-tailed deer is one of the most adaptable animals in the world. Their range extends from the southern tip of Hudson Bay in Canada well into South America. They live in areas where climatic factors vary from humid, tropical jungle to dry, hot desert to northern subarctic conditions. This species has probably extended its range northward as a result of forestry and agricultural practices. White-tails are very tolerant of people and their practices. A small subspecies of white-tailed deer, the Key deer (*O. v. clavium*), occurs in the Florida Keys and is classified as endangered. The mule deer inhabits most of temperate North America between the Pacific coast and the 100th meridian. Mule deer may be distinguished from white-tailed deer by a number of characteristics, including the rather long black-tipped tail, the long ears, the shape and position of metatarsal glands, the forked antlers having two tines of equal size and branching equally above the base to form four major tines, and their behavior. Antlers of adult white-tailed deer consist of a series of tines arising from a continuous, forecurving main beam. The elk or wapiti is a group of subspecies of the red deer complex that is distributed across North America. The entire species is distributed across Eurasia as well as North America. Elk once occurred in virtually all of temperate North America, excluding the Great Basin in Nevada and the southeastern United States. Merriam's elk and the eastern elk are now extinct, and the prairie and deciduous hardwood portions of the original range are no longer occupied, except for small, isolated introduced populations. Moose, the largest living members of the deer family, are found from the Rocky Mountain region to Maine and north through Canada and Alaska. The long legs and short neck make it impossible for a moose to reach short grass unless it kneels; in order to drink, it must wade into the water. The names caribou and reindeer refer to the same species: in the wild state these animals are known as caribou, whereas the domesticated form is known as reindeer. With the exception of the musk ox, caribou live farther north than any other hoofed animal. They range beyond the timber line into the Arctic Circle in both the New World and the Old World. *See* ARTIODACTYLA; CARIBOU; MOOSE.

Donald W. Linzey

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Definite composition, law of

The law that a given chemical compound always contains the same elements in the same fixed proportions by weight. Thus, whatever its source, silver chloride always contains 108 g (3.81 oz) of silver to every 35.45 g (1.251 oz) of chlorine. If a compound is formed by the union of m atoms of one element, each weighing a, with n atoms of another element, each weighing b, the composition by weight of one molecule of the compound is in the ratio ma:nb. This must be the composition of any mass of the

compound, provided that all atoms of the same kind have the same weight. It is now known that this is not usually the case but that the atoms of an element may consist of a number of isotopes, having different masses. However, as long as any sample of the element always contains the same relative proportions of the isotopes, the law still holds. *See* ATOMIC MASS; ISOTOPE.

This is not the case for lead. Lead is the final product of the decay of three radioactive series, the atomic weights from the three series being 206 from radium, 208 from thorium, and 207 from actinium. Hence both the atomic weight of lead and the proportion of lead in its compounds will vary with the source of the lead. *See* RADIOACTIVITY.

Much more widespread and serious departures from the law of definite composition occur in a large variety of solid compounds (the nonstoichiometric compounds). Nonstoichiometry arises for a number of differing reasons. In the silicate minerals, such as olivine, it occurs as a result of isomorphous replacement. Thus olivine is $(Mg,Fe)_2SiO_4$ and the proportion of magnesium to iron may vary widely from sample to sample. Other solids are simply deficient in metal atoms. Thus ferrous sulfide, FeS, rarely has an atomic ratio Fe/S of precisely unity. Density measurements have shown that the lattice of the sulfur atoms remains intact but that some of the iron atoms are missing. *See* CRYSTAL STRUCTURE; NONSTOICHIOMETRIC COMPOUNDS; SILICATE MINERALS; STOICHIOMETRY.

Thomas C. Waddington

Defoliant and desiccant

Defoliants are chemicals that cause leaves to drop from plants; defoliation facilitates harvesting. Desiccants are chemicals that kill leaves of plants; the leaves may either drop off or remain attached; in the harvesting process the leaves are usually shattered and blown away from the harvested material. Defoliants are desirable for use on cotton plants because dry leaves are difficult to remove from the cotton fibers. Desiccants are used on many seed crops to hasten harvest; the leaves are cleaned from the seed in harvesting.

True defoliation results from the formation of an abscission layer at the base of the petiole of the leaf. Most of the chemicals bring about this type of defoliation, and the leaves abscise and drop from the plant. Certain chemicals kill plant leaves at low application rates, resulting in desiccation but not abscission. *See* ABSCISSION.

The most common agency of defoliation in nature is frost; frosted leaves of cotton dry up and fall off after a few days. With the rapid increase in mechanical harvesting of crops, new chemicals and new processes have been introduced; these are termed harvest aids. Most harvest-aid chemicals are growth regulators that function at low concentrations or at low application rates; in some way they stimulate or inhibit growth. Some serve to hold fruits such as apples and pears upon the trees, thus preventing preharvest drop. Others cause loosening of fruits, facilitating harvest by shaking; examples are sour cherries and grapes.

Because of the advantages of harvest aids, much effort has gone into the search for new chemicals.

A list of active ingredients in some harvest aids includes (trade names are given in parentheses) ametryn (Evik), amino triazole, ammonia gas, ammonium nitrate, ammonium thiocyanate, arsenic acid, cacodylic acid, calcium cyanamide, diethyl dithio-bisthionoformate, diquat, endothal (Accelerate, Des-I-Cate), ethephon (Ethrel), 4,6-dinitro-*o*-*sec*-butylphenos, hexachloroacetone, magnesium chlorate, nitrophen, oxyfluorfen, paraquat, pentachlorophenol, petroleum solvents, sodium borate, sodium cacodylate (Bollseye), sodium chlorate (Tumbleleaf), sodium *cis*-3-chloroacrylate, sodium naphthalene acetate (Fruitone N), tributylphosphorothioate (Folex), tributylphosphorotrithioate (Def-6), 2,4,5-TP (Silvex), triethanolamine salt of Silvex (Fruitone T).

Crops upon which harvest aids are used include alfalfa, apples, blackberries, blueberries, clovers (Alsike, Ladino), cantaloupes, castor beans, cherries, cranberries, flax, figs, filbert nuts, grapes, guar, hops, lemons, onions, oranges, peas, potatoes, rice, safflower, sudan grass, sorghum, soybeans, sunflowers, tangerines, tomatoes, walnuts, and wheat. Tomato plants are sometimes cut below the soil surface and allowed to dry out; this brings the fruits to a uniform ripeness for mechanical harvesting.

Modern agricultural technology, which has greatly increased production in the United States, is dependent upon harvest-aid chemicals. These aids bring fruits to uniform ripeness for harvest by machinery or by well-managed crews. They reduce or eliminate much labor, save time, and make for reliable management of harvest. They, along with fertilizers and improved crop varieties, are responsible for the great abundance of foods in developed countries.

It was in an effort to find such chemical agents that 2,4-D was discovered in 1942 by E. J. Kraus. For decades 2,4-D and 2,4,5-T have been important compounds used for weed control.

Defoliants and desiccants have also been used during war to destroy crops. *See* HERBICIDE.

Alden S. Crafts

Degeneracy (quantum mechanics)

A term referring to the fact that two or more stationary states of the same quantum-mechanical system may have the same energy even though their wave functions are not the same. In this case the common energy level of the stationary states is degenerate. The statistical weight of the level is proportional to the order of degeneracy, that is, to the number of states with the same energy; this number

is predicted from Schrödinger's equation. The energy levels of isolated systems (that is, systems with no external fields present) comprising an odd number of so-called fermions (for example, electrons, protons, and neutrons) always are at least twofold degenerate. *See* EXCLUSION PRINCIPLE.

Relation to symmetries. Except for so-called accidental degeneracy, degeneracy is associated with special symmetries of the physical system, and can be removed by destroying this symmetry. For example, an energy level of total angular momentum J (in units of $(\hbar/2\pi$, where $\hbar$ is Planck's constant) has a $(2J + 1)$-fold degeneracy that results from the circumstance that the $2J + 1$ allowed values of J_z, the z component of total angular momentum, all have the same energy. In a magnetic field H along z, this level splits into $2J + 1$ energy levels, since the energy now depends on J_z, that is, on the angle between H and the total angular momentum. *See* ANGULAR MOMENTUM.

Relation to perturbation theory. In applications of quantum mechanics to complicated systems, such as many-electron atoms (in atomic physics applications) or heavy atomic nuclei (in nuclear physics applications), the ability to associate degenerate or nearly degenerate levels with underlying symmetries is crucial to successful calculations of the needed stationary-state energies and wave functions via perturbation theory, which often is the most convenient calculational method for such systems. In stationary-state perturbation theory the full hamiltonian H is written in the form $H = H_0 + V$, where H_0 is regarded as the unperturbed hamiltonian and V is the perturbing interaction, which, presumably, is comparatively small. Typically, however, the actual "smallness'' of the perturbation—and therefore the validity of the perturbation theory for computing the energy and wave function of a stationary state i of H starting from the corresponding unperturbed energy (also indexed by i) and wave function—is measured by ratios of the form $V_{ij}/(E_i - E_j)$, where j is another stationary state of the unperturbed hamiltonian, V_{ij} is the matrix element of V between the unperturbed states i and j, and E_i and E_j denote respectively the unperturbed energies of the states i and j. Consequently, stationary-state perturbation theory breaks down when there are states for which $E_i - E_j = 0$, that is, when the energy shift from the unperturbed energy is sought for a state i belonging to a group of degenerate unperturbed states, unless the original set of unperturbed states are chosen so that every matrix element $V_{ij} = 0$ whenever $E_i - E_j = 0$. In the magnetic field illustration discussed above, for instance, in order to calculate the level splitting in heavy atoms (where the total orbital angular momentum and the total spin generally no longer are good quantum numbers) by perturbation theory, it is necessary to choose the $2J + 1$ unperturbed originally degenerate states so that J_z is a good (and different) quantum number in each of those unperturbed states; in other words, each of those $2J + 1$ originally degenerate states must be an eigenstate belonging to a different value of J_z in the allowed range $-J \leq J_z \leq J$.

For a state i of the full hamiltonian corresponding to a nondegenerate state of the unperturbed hamiltonian, choosing a set of unperturbed states for which every $V_{ij} = 0$ is not a prerequisite for use of stationary perturbation theory. The above-stated measure of the validity of stationary-state perturbation theory suggests, however, and careful analysis confirms, that the energy shift of a perturbed state i from its unperturbed value (whether degenerate or nondegenerate) generally is much more strongly influenced by its "interaction" V_{ij} with any state j whose unperturbed energy E_j lies close to E_i than by interactions with states whose unperturbed energies are quite different from E_i unless V_{ij} is zero or exceptionally small. Similarly, unless V_{ij} happens to be zero or exceptionally small, the wave function of the perturbed state i (whether degenerate or nondegenerate) will be much more strongly "mixed'' with the unperturbed wave function of any state j having a nearby unperturbed energy E_j than with the wave functions of states whose unperturbed energies are quite different from E_i.

In quantum mechanics and in other branches of mathematical physics, the term degeneracy is employed also to characterize the eigenvalues of operators other than the energy operator. *See* EIGENVALUE (QUANTUM MECHANICS); NONRELATIVISTIC QUANTUM THEORY; PERTURBATION (QUANTUM MECHANICS); SYMMETRY LAWS (PHYSICS). Edward Gerjuoy

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Degree-day

A unit used in estimating energy requirements for building heating and, to a lesser extent, for building cooling. It is applied to all fuels, district heating, and electric heating. Origin of the degree-day was based on studies of residential gas heating systems. These studies indicated that there existed a straight-line relation between gas used and the extent to which the daily mean outside temperature fell below 65°F (18°C).

The number of degree-days to be recorded on any given day is obtained by averaging the daily maximum and minimum outside temperatures to obtain the daily mean temperature. This procedure was found to be adequate and less time consuming than the averaging of 24 hourly temperature readings. The daily mean so obtained is subtracted from 65°F (18°C) and tabulated. Monthly and seasonal totals of degree-days are available from local weather bureaus. They use the 30-year period 1931–1960 as their basis.

The base temperature of 65°F (18°C) was established by observation of the outside temperatures at which the majority of homeowners started their gas heating systems in the early fall. Interpreted in another way, the internal heat gains in residences,

which appear to be about 5000 Btu/h (1500 W), are sufficient to provide heat down to a balance temperature of 65°F (18°C).

Provided that the efficiency of the heating equipment remains constant, the fuel usage of a given house will be proportional to the number of degree-days. For purposes of comparing one heating season with another, constant efficiency is a reasonable assumption. However, efficiency of the heating equipment decreases markedly in the warmer portions of the heating season. Thus the fuel use per degree-day will be higher in spring and fall than in winter. The same considerations apply to use of degree-day data in warmer climates compared to cooler regions.

A frequent use of degree-days for a specific building is to determine before fuel storage tanks run dry when fuel oil deliveries should be made.

Number of Btu which the heating plant must furnish to a building in a given period of time is Btu required = heat rate of building × 24 × degree-days where "Btu required" is the heat supplied by the heating system to maintain the desired inside temperature. "Heat rate of building" is the hourly building heat loss divided by the difference between inside and outside design temperatures.

To determine fuel requirement from this relationship, two decisions are needed. First, the correct seasonal efficiency of the heating system must be obtained or estimated; second, appropriate base temperature for the degree-day data must be calculated. Seasonal efficiency is generally less than the test efficiency of the heating unit although the load efficiency curves are surprisingly level down to quite low loads. Base temperature for degree-day values may be chosen from consideration of internal heat gains, as in a well-lighted office building, then degree-day data on other than a 65°F (18°C) basis are required. Balance temperatures for both day operation and night operation may also need to be taken into consideration. Then it would be proper to replace the term degree-days in the equation with the term adjusted degree-days. *See* AIR CONDITIONING; CLIMATOLOGY; COMFORT HEATING; COMFORT TEMPERATURES; PSYCHROMETRICS.

C. George Segeler

Degree of freedom (mechanics)

Any one of the number of independent ways in which the space configuration of a mechanical system may change. A material particle confined to a line in space can be displaced only along the line, and therefore has one degree of freedom. A particle confined to a surface can be displaced in two perpendicular directions and accordingly has two degrees of freedom. A particle free in physical space has three degrees of freedom corresponding to three possible perpendicular displacements. A system composed of two free particles has six degrees of freedom, and one composed of N free particles has $3N$ degrees. If a system of two particles is subject to a requirement that the particles remain a constant distance apart, the number of degrees of freedom becomes five. Any requirement which diminishes by one the degrees of freedom of a system is called a holonomic constraint. Each such constraint is expressible by an equation of condition which relates the system's coordinates to a constant, and may also involve the time. When applied to systems of particles, a holonomic constraint frequently has the geometrical significance of confining a particle to a specified surface, which may be time-dependent. *See* CONSTRAINT.

A mechanical system may be conceived as a set of N particles in space subject to a certain number K of holonomic constraints which reduce the number of degrees of freedom to $3N - K$. In the case of a rigid structure consisting of more than two noncollinear particles, the number of degrees of freedom is independent of the number of particles and is equal to six.

Russell A. Fisher

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De Haas-van Alphen effect

An oscillatory behavior of the magnetic moment of a pure metal crystal with changes in the applied magnetic field B, at very low temperatures. It is named after its discoverers W. J. de Haas and P. M. van Alphen. This effect has its origin in the Bohr-Sommerfeld quantization of the orbits of conduction electrons under the influence of the magnetic field.

Origin. A free electron in a magnetic field B moves in a circular orbit with an angular frequency given by the cyclotron resonance frequency, Eq. (1), where e

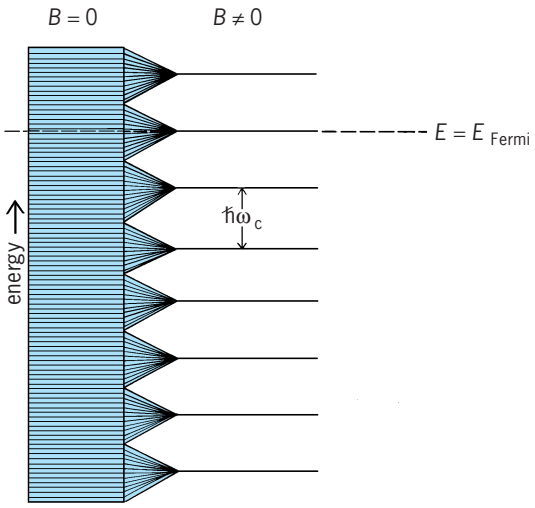
$$\omega_c = \frac{eB}{m} \tag{1}$$

and m are the charge and mass of the electron. But the energy E of this simple harmonic motion must be quantized according to the Planck prescription, Eq. (2), where $\hbar$ is Planck's constant divided by 2π,

$$E_n = (n + {}^1/_2)\hbar\omega_c \tag{2}$$

and n numbers the quantized energy levels.

In a metal at low temperatures at zero magnetic field, electrons are not at rest, but have a continuous distribution of energies up to some limiting energy E_F called the Fermi energy; this arises because the Pauli exclusion principle keeps the electrons from all falling into the lowest energy state. Those electrons with energy close to E_F dominate the behavior of the metal. In a magnetic field the allowed electron energies are not continuous, but must correspond to one of those given by Eq. (2). Suppose for some value of n, B_n an electron energy level matches



Distribution of electron energy in a metal at zero and finite magnetic field *B*, showing the condition that an electron energy level matches the Fermi energy.

the Fermi energy E_F, that is, $E_n = E_F$ (see **illus.**); from Eqs. (1) and (2), this corresponds to Eq. (3).

$$E_F = \frac{(n + {}^1/_2)\hbar e}{m} B_n \tag{3}$$

See CYCLOTRON RESONANCE EXPERIMENTS; EXCLUSION PRINCIPLE; FREE-ELECTRON THEORY OF METALS; NONRELATIVISTIC QUANTUM THEORY.

In order for n to increase by 1, the magnetic field must change so that Eq. (4) is satisfied. Eliminating n between Eqs. (3) and (4) yields Eq. (5) for the amount

$$E_F = \frac{(n + {}^3/_2)\hbar e}{m} B_{n+1} \tag{4}$$

$$\Delta\left(\frac{1}{B}\right) \equiv \frac{1}{B_{n+1}} - \frac{1}{B_n} = \frac{e\hbar}{mE_F} \tag{5}$$

by which the magnetic field must change in order for n to increase by 1. The quantity $\Delta(1/B)$ determines the "period" of the de Haas–van Alphen oscillations (it is a period in magnetic field, not time). Since $E = p^2/2m$, where p is the electron momentum, Eq. (6)

$$\Delta\left(\frac{1}{B}\right) = \frac{2e\hbar}{p_F^2} \tag{6}$$

is valid, where p_F is the limiting momentum corresponding to the energy E_F. Thus, a measurement of the period of the oscillations permits a determination of the limiting or Fermi momentum of the electrons. A detailed quantum statistical-mechanical treatment shows that many of the properties of a metal, including the magnetization, oscillate with field according to Eq. (6). *See* MAGNETIZATION.

Measurement of effective mass. The assumption that the electrons occupy only states with E_n equal to or less than E_F involves the implicit assumption that the metal is cooled to absolute zero. At any finite temperature T, states with E_n greater than E_F will also be occupied. The ratio of the number of electrons in two neighboring energy levels E_n and E_{n+1} is governed by the Boltzmann occupation factor $\exp[(E_{n+1} - E_n)/k_BT]$, where k_B is Boltzmann's constant. If T is large enough such that k_BT is much greater than $\hbar\omega_c$, this factor becomes unity, and in this case there is nothing particularly significant about the condition $E_F = E_n$; that is, the amplitude of the oscillations is expected to be greatly attenuated. On the other hand, at very low temperatures where $\hbar\omega_c$ is greater than k_BT, strong oscillations are expected. Thus, a measurement of the temperature dependence of the oscillation amplitude permits a determination of the cyclotron frequency, or equivalently the electron mass. In a metal, as in a semiconductor, electrons behave with an effective mass m^* rather than the free electron mass, and the de Haas–van Alphen effect thus allows a measurement of m^* in metals.

Determination of Fermi surface. When electrons move in a metal, the energy-momentum relation in general differs radically from that of a free particle. The momenta $\vec{p}_F$ which correspond to the Fermi energy depend on the direction $\hat{p}_F$ within the crystal; that is, the shape of the Fermi surface differs from a sphere.

Application of the Bohr-Sommerfeld quantization rules shows that Eq. (6) for the de Haas–van Alphen period must be generalized to Eq. (7), where A_p is

$$\Delta\left(\frac{1}{B}\right) = \frac{2\pi\hbar e}{A_p} \tag{7}$$

the area of the orbit in momentum space, that is, the cross section of the Fermi surface.

For a given Fermi surface, the cross-sectional area is a function of the momentum parallel to the field and ranges between various extremal values. In principle, all of these sections make an oscillatory contribution. However, for all sections other than extremal sections, the oscillatory contributions at neighboring sections tend to cancel. Thus in Eq. (7) A_p is the area of an extremal section (or sections for metals whose Fermi surface supports multiple extremal cross sections). By studying the dependence of A_p on the direction B of the magnetic field, sufficient information can be obtained to construct the detailed shape of the Fermi surfaces of a metal. A further condition for observation, in addition to low temperatures, is sample purity: the electron-scattering time τ must be longer than, or at least comparable to, a cyclotron period $2\pi/\omega_c$. *See* RELAXATION TIME OF ELECTRONS.

Virtually all metals available in sufficient purity have been studied by the de Haas–van Alphen effect. It is also the most powerful probe of Fermi surface properties in alloys and intermetallic compounds.

The intermetallic compounds of the greatest interest are the so-called heavy-fermion materials involving actinide and rare-earth atoms. Effective masses greater than 1000 have been seen in electronic

heat-capacity measurements. In one of these materials, UPt_3, a mass of 200 electron masses was observed, the largest value ever measured by using the de Hass–van Alphen effect.

Prescriptions for deconvolution of extremal areas to Fermi momenta are well developed. Furthermore, measurements of the angular-dependent effective mass m^* can be used to determine the velocity of electrons at the Fermi surface. *See* BAND THEORY OF SOLIDS; FERMI SURFACE.

J. B. Ketterson

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Dehumidifier

Equipment designed to reduce the amount of water vapor in the atmosphere. *See* HUMIDITY.

The atmosphere is a mechanical mixture of dry air and water vapor, the amount of water vapor being limited by air temperature. Water vapor is measured in either grains per pound of dry air or pounds per pound of dry air (7000 gr = 1 lb).

There are three methods by which water vapor may be removed: sorbent materials; cooling to the required dew point; and compression with aftercooling.

Sorbent type. Sorbents are materials which are hygroscopic to water vapor; they are available in both solid and liquid forms. Solid sorbents include silica gels, activated alumina, and aluminum bauxite. Liquid sorbents include halogen salts such as lithium chloride, lithium bromide, and calcium chloride, and organic liquids such as ethylene, diethylene, and triethylene glycols and glycol derivatives.

Solid sorbents may be used in static or dynamic dehumidifiers. Bags of solid sorbent materials within packages of machine tools, electronic equipment, and other valuable materials subject to moisture damage constitute static dehumidifiers. An indicator chemical may be included to show by a change in color when the sorbent is saturated. The sorbent then requires reactivation by heating at 300–350°F (149–177°C) for 1–2 h before reuse.

A dynamic dehumidifier for solid sorbent consists of a main circulating fan, one or more beds of sorbent material, reactivation air fan, heater, mechanism to change from dehumidifying to reactivation, and aftercooler.

A single-bed dehumidifier (**Fig. 1**) operates on an intermittent cycle of dehumidifying for 2–3 h and then switches to the reactivation cycle for 15–45 min. No dehumidification is obtained during the reactivation cycle. A single-bed unit is used for small areas where moisture is a problem. The moist reactivation air is discharged to the outside.

The dual-bed machine is larger in capacity than the single-bed unit and has the advantage of providing a continuous supply of dehumidified air (**Fig. 2**). While one bed is dehumidifying, the other bed is reactivating. After a preset time interval, the air cycle is switched to pass the air through the reactivated bed for dehumidification and to reactivate the saturated bed.

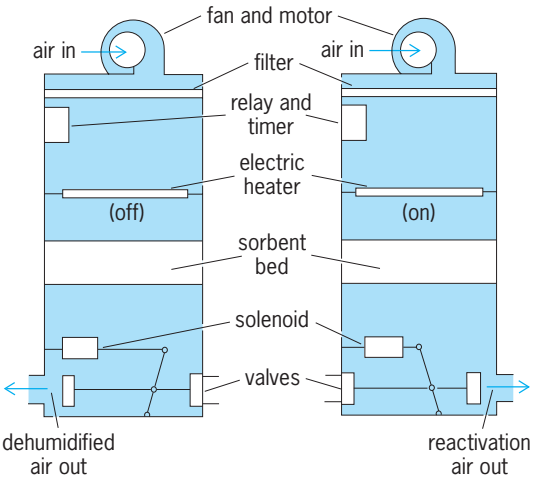

Fig. 1. Single-bed solid-sorbent dehumidifier. Dehumidifying cycle on left and reactivation cycle on right.

The dew point of the effluent air of a fixed-bed machine is lowest at the start of a cycle immediately after the reactivated bed has been placed in service. The dew point gradually rises as the bed absorbs the water vapor and eventually would be the same as the entering dew point when the vapor pressure of the sorbent reached the vapor pressure of the air and could no longer absorb moisture from the air. The cutoff point at which the absorbing bed is changed over to reactivation is fixed by the maximum allowable effluent dew point. *See* DEW POINT.

A multibed unit with short operating cycles will reduce the range of effluent dew point to within a few degrees. A multibed unit with rotating cylindrical bed maintains a reasonably constant effluent dew point.

The liquid-sorbent dehumidifier consists of a main circulating fan, sorbent-air contactor, sorbent pump,

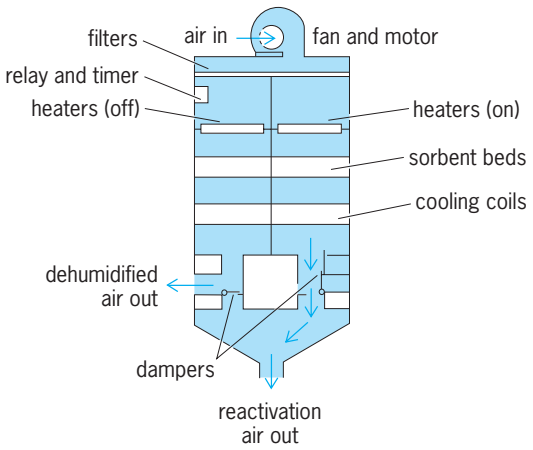

Fig. 2. Dual-bed solid-sorbent dehumidifier. Air is being dehumidified through the left bed at the same time that the right bed is being reactivated.

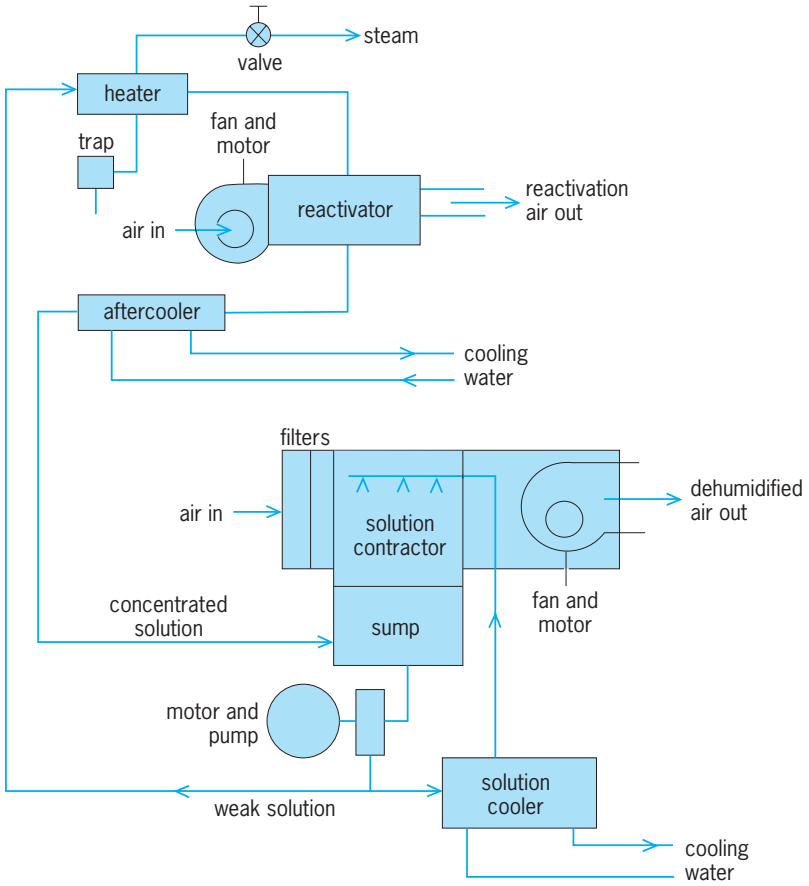

Fig. 3. Liquid-sorbent dehumidifier with continuous dehumidifying and reactivation.

and reactivator including contactor, fan, heater, and cooler (**Fig. 3**). This unit will control the effluent dew point at a constant level because dehumidification and reactivation are continuous operations with a small part of the sorbent constantly bled off from the main circulating system and reactivated to the concentration required for the desired effluent dew point.

Cooling type. A system employing the use of cooling for dehumidifying consists of a circulating fan and cooling coil. The cooling coil may use cold water obtained from wells or a refrigeration plant, or may be a direct-expansion refrigeration coil. In place of a coil, a spray washer may be used in which the air passes through two or more banks of sprays of cold water or brine, depending upon the dew-point temperature required.

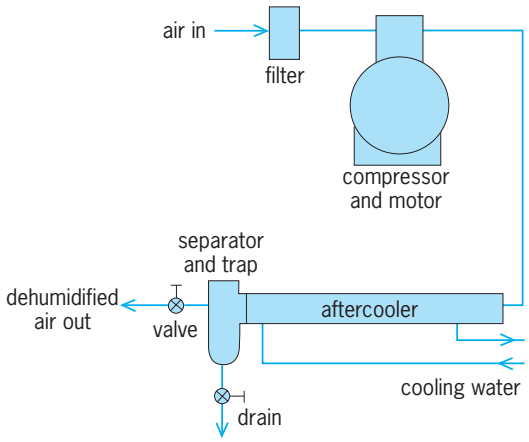

Fig. 4. Dehumidifying by compression and aftercooling.

When coils are used, the leaving dew point is seldom below 35°F (1.7°C) because of possible buildup of ice on the coil. When it is necessary to use coils for temperatures below 35°F (1.7°C), as in cold-storage rooms, either two coils are used so one can be defrosted while the other is in operation, or only one coil is used and dehumidifying is stopped during the defrost period.

A brine-spray dehumidifier or brine-sprayed coil can produce dew-point temperatures below 35°F (1.7°C) without frosting if properly operated and maintained. *See* AIR COOLING.

Compression type. Dehumidifying by compression and aftercooling is used when the reduction of water vapor in a compressed-air system is required. This is particularly important, for example, if the air is used for automatic control instruments or cleaning of delicate machined parts.

If air is compressed and the heat of compression removed to bring the temperature of the air back to the temperature entering the compressor, condensation will take place and the remaining water vapor content will be directly proportional to the absolute pressure ratio of the compressed air (**Fig. 4**).

For example, if saturated air at 70°F or 39°C (111 gr/lb of dry air) is compressed from atmospheric pressure (14.7 lb/in.^{2} absolute) to 88 lb/in.^{2} absolute (6:1 compression ratio) and cooled to 70°F (39°C), the remaining water vapor in the compressed air will be 111/6 = 18.5 gr/lb of dry air. If the air is expanded back to atmospheric pressure and 70°F (39°C), the dew point will be 24°F (13°C). *See* PSYCHROMETRICS.

The power required for compression systems is so high compared to power requirements for dehumidifying by either the sorbent or refrigeration method that the compression system is not an economical one if dehumidifying is the only end result required. *See* COMPRESSOR; DRYING.

John Everetts, Jr.

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Dehydrogenation

A reaction in which hydrogen is detached from a molecule. The reaction is strongly endothermic; therefore, heat must be supplied to maintain the reaction temperature. Since carbon-carbon bonds are easier to break than carbon-hydrogen bonds, highly selective reaction conditions are required to cause the molecule to dehydrogenate rather than break apart into small molecules (for example, methane) and carbon (coke). Superheated steam (725°C or 1337°F) is

often mixed with the dehydrogenation reactor feed to provide the needed heat of reaction, lower the partial pressure and enhance the equilibrium conversion, and decoke the catalyst to extend the reactor onstream time. In addition, excess hydrogen is sometimes added to the feed to depress coke formation and diminish the breakup of large molecules (for example, styrene) into smaller molecules of much less economic value.

Industrial applications. Dehydrogenation reactions are of major industrial importance. They account for the production of tremendous quantities of organic compounds that enter into the manufacture of textiles, gasoline, lubricants, explosives, plastics, plasticizers, agricultural chemicals, perfumes, detergents, elastomers, and pharmaceutical products.

Most commercial dehydrogenation reactions are catalytic and take place in the vapor phase. However, the noncatalytic, thermal dehydrogenation of isobutane is practiced to produce propylene and isobutylene. Also, alcohols with high boiling points can be dehydrogenated in the liquid phase in the presence of powdered catalysts (for example, active Raney nickel). The primary catalysts used for dehydrogenation reactions are the transition metals: Ti, V, Cr, Mn, Fe, Co, Ni, Cu, Zn, Pd, W, and Pt. *See* CATALYSIS; HETEROGENEOUS CATALYSIS.

Reactions. The primary types of dehydrogenation reactions are:

1. Vapor-phase conversion of primary alcohols to aldehydes, as in the preparation of acetaldehyde from ethyl alcohol in the presence of a silver catalyst at 550°C (1020°F), reaction (1).

$$CH_3CH_2OH + {}^1/_2O_2 \longrightarrow CH_3CHO + H_2O \quad (1)$$

2. Vapor-phase conversion of secondary alcohols to ketones, as in the preparation of acetone from isopropyl alcohol, reaction (2), and methyl ethyl ketone from secondary alcohol in the presence of copper catalyst at 300°C (570°F), reaction (3).

$$CH_3CHOHCH_3 \longrightarrow CH_3COCH_3 + H_2 \quad (2)$$

$$CH_3CHOHCH_2CH_3 \longrightarrow CH_3COCH_2CH_3 + H_2 \quad (3)$$

3. Dehydrogenation of paraffins to monoolefins, as in the Houdry Catofin process to manufacture propylene from propane, in the presence of chromia-alumina catalyst at 550–600°C (1020–1110°F), reaction (4).

$$CH_3CH_2CH_3 \longrightarrow CH_2{=}CHCH_3 + H_2 \quad (4)$$

4. Conversion of monoolefins to diolefins, practiced in the conversion of 1-butene to 1,3-butadiene, over chromia-alumina or noble-metal catalysts at 550–600°C (1020–1110°F), reaction (5).

$$CH_2{=}CHCH_2CH_3 \longrightarrow CH_2{=}CHCH{=}CH_2 + H_2 \quad (5)$$

5. Dehydrogenation of a side chain, commercialized in the Monsanto/Lummus and Cosden/Badger processes for the manufacture of styrene from ethylbenzene in the presence of iron oxide catalyst at 600°C (1110°F), reaction (6).

$$C_6H_5C_2H_5 \longrightarrow C_6H_5CH{=}CH_2 + H_2 \quad (6)$$

6. Manufacture of aromatics by the dehydrogenation of cycloparaffins, as in the catalytic reforming of naphthas and naphthenes in the presence of a platinum catalyst at 500°C (930°F) and 100–1000 lb/in.^{2} (700–7000 kilopascals) for the production of toluene and high-octane gasoline.

Oxidative dehydrogenation. When the detached hydrogen from the dehydrogenation reaction is immediately oxidized, two benefits accrue: (1) the conversion of reactants to products is increased because the equilibrium concentration is shifted toward the products (law of mass action), and (2) the added exothermic oxidation reaction supplies the needed heat of reaction. An example is the oxidative dehydrogenation of 1-butene to 1,3-butadiene practiced in the PetroTex Oxo-D process, reaction (7).

$$CH_2{=}CHCH_2CH_3 + {}^1/_2O_2 \longrightarrow CH_2{=}CHCH{=}CH_2 + H_2O \quad (7)$$

An innovative concept for the oxidative dehydrogenation of paraffins, which has only been demonstrated experimentally, is the use of molten halide catalysts on porous supports. This type of catalyst uses a binary eutectic mixture of alkali metal halides (for example, KCl-LiCl) with added transition-metal promoter (for example, Mn) on a low-surface-area support (for example, 15 m^2/g alumina or 4500 ft^2/oz). Under reaction conditions, the catalytic material exists as a thin, molten film which covers the inner surface area of the support. (This is similar to the sulfuric acid catalyst, which exists as a vanadium pyrosulfate melt lining the porous pellets during reaction.)

Oxidative dehydrogenation reaction systems suffer one major drawback: the added oxygen will react with the hydrocarbon as well as the hydrogen, generating carbon dioxide and other undesired oxides. Unless the oxygen-hydrocarbon reactions can be depressed, an oxidative dehydrogenation system will not be economical.

Dual-function catalysts. Dual-function catalysts have been developed which promote both dehydrogenation and a subsequent reaction, to reduce the total reactor volume and processing steps required to manufacture the desired end product from the raw material. Generally, the catalytic agent (for example, Pt) for dehydrogenation is bound to a support which has acid sites (for example, halogen-activated alumina). One application is the dehydrogenation and isomerization of naphthenes in catalytic reforming. Another example is dehydrocyclodimerization, the conversion of small (C_3–C_5) paraffins or olefins to aromatics. Xylenes, ethylbenzene, toluene, and benzene are produced from butanes by this process.

Biochemical dehydrogenation. Dehydrogenation reactions are essential steps in the metabolism and

biosynthesis of living organisms. The catalysts for these reactions are dehydrogenases, which catalyze the removal of hydrogen atoms from fuel molecules and their transfer to coenzymes. These coenzymes, in turn, carry the energy-rich electrons from catabolic reactions to electron-requiring biosynthesis reactions. *See* CHEMICAL CONVERSION; ENZYME; HYDROGENATION; METABOLISM; OXIDATION PROCESS. J. W. Fulton

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Delay line

A transmission line (as nearly dissipationless as possible) or an electric network approximation of it which, if terminated in its characteristic impedance, will reproduce at its output a waveform applied to its input terminals with little distortion but at a time delayed by an amount dependent upon the electrical length of the line.

If a transmission line is dissipationless, which will be the case if its series resistance is zero and its shunt conductance is also zero, it will have a characteristic impedance Z_0, as shown in Eq. (1), where L is the

$$Z_0 = \sqrt{L/C} \tag{1}$$

series inductance and C the shunt capacitance per unit length of the line. *See* TRANSMISSION LINES.

The velocity v of propagation of a signal along the line is shown by Eq. (2).

$$v = \frac{1}{\sqrt{LC}} \tag{2}$$

Therefore the time required for the pulse to propagate a distance x along the line is shown by Eq. (3). Such a line, terminated in its characteristic

$$T_d = x\sqrt{LC} \tag{3}$$

impedance, is shown in **Fig. 1**; and output pulses reproduce input at a delayed time T_d.

The lumped-circuit approximation of the transmission line is shown in **Fig. 2**. If the inductance and capacitance per section are L_1 and C_1, then the total time delay is shown by Eq. (4), where n is the number of sections.

$$T_d = n\sqrt{L_1C_1} \tag{4}$$

If the delay line is not terminated in its characteristic impedance, there is multiple reflection back and forth along the line. For example, if the receiving end of the line is unterminated, as in **Fig. 3**, but the sending end is terminated in its characteristic impedance, the receiving-end reflection coefficient is positive, and a delayed pulse appears at the output, as given by Eqs. (3) and (4). In addition, a pulse

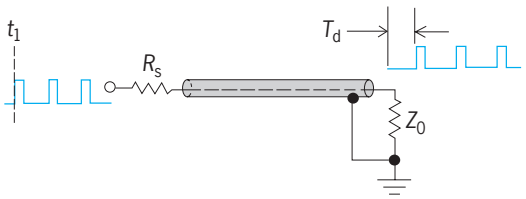

Fig. 1. Transmission line as delay line.

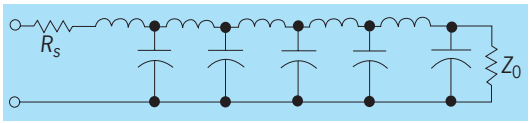

Fig. 2. Lumped-circuit delay line.

of the same polarity appears at the input terminals, delayed by twice this amount (the time required for a pulse to travel to the end and back).

If the receiving end is terminated in a short circuit as shown in **Fig. 4**, the receiving-end reflection coefficient is negative, and no pulse appears at the output, although an inverted pulse will appear at the input terminals with the same time delay as before.

Various applications make use of the short circuit and open circuit of delay lines, including line-controlled pulse generators. *See* PULSE GENERATOR.

Delay lines are also used for establishing a time sequence for the occurrence of events. A delay line with a total length equal to the greatest time delay required in a system may be used as a basic element. Pulses occurring at intermediate times may be obtained from taps at various points along the line. A specific application is found in the synchronizing signal generator of the television system. Also, the lumped-circuit delay line is an essential element of the wide-band distributed amplifier.

When a signal is digital in nature, or consists of a series of pulses, the series of pulses may be delayed by using a shift register, which might, for

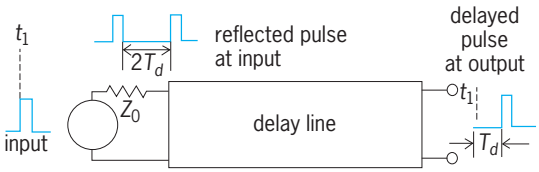

Fig. 3. Reflection due to unterminated receiving end.

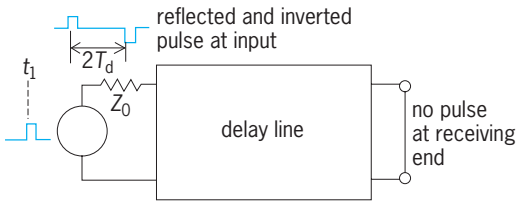

Fig. 4. Reflection due to short-circuited receiving end.

example, consist of a chain of cascaded type D flip-flops. If the register has n stages, the pulse series will appear at the output delayed by a time $(n - 1)T$, where T is the interval between consecutive pulses of the system timing clock. The same function can be realized by using switched-capacitor circuits or an array of charge-coupled devices. *See* CHARGE-COUPLED DEVICES; SWITCHED CAPACITOR; TRANSMISSION LINES.

Glenn M. Glasford

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Delayed neutron

A neutron emitted spontaneously from a nucleus as a consequence of excitation remaining from a preceding radioactive decay event. Analogously, delayed emission of protons and alpha particles is observed, but the known delayed-neutron emitters are more numerous, and some of them have practical implications. In particular, they are important in the control of nuclear chain reactors. *See* CHAIN REACTION (PHYSICS); NUCLEAR REACTOR.

The delayed neutrons were discovered in 1939, just after the discovery of fission. The first measure of their abundance and the first resolution of their decay periods were made in 1942 and were instrumental in the decision to attempt the first nuclear chain reaction in the midst of a large city. The first chemical identifications of the delayed-neutron emitters were made in 1942 and published in 1945.

Nuclear origin. Radioactive transformation by beta decay often leaves a product nucleus with internal energy in excess of that associated with its stable state, and usually the energy is radiated as gamma-ray quanta. In exceptional cases, the energy may exceed that required to remove one or more of the constituent neutrons from the product nucleus; neutron emission then rapidly takes place and at least partially deexcites the nucleus. The delayed neutrons are accordingly a radioactive emission, occurring essentially at the time of decay of the beta-emitting precursor. If the precursor is formed, for example, by fission, the delayed neutrons may be emitted seconds, or even minutes, after the moment of fission (because of the intervening beta decay), and are to be distinguished from the prompt neutrons that are emitted at the time of fission. On the other hand, the delayed neutrons are emitted so quickly after the completion of the beta decay that the half-lives and the chemical behavior of the delayed-neutron radioactivities are, in practice, those of the precursors.

Occurrence. Delayed-neutron emission occurs in neutron-rich nuclei, that is, in the heavier isotopes of various elements. These isotopes have several neutrons in excess of the proton-neutron mixture associated with the element's stable isotopes; in consequence, their beta-decay half-lives are short. In relation to the shell model of the nucleus, they are likely to be associated also with nuclides having one or a few neutrons in excess of a closed neutron shell, because such neutrons are expected to be loosely bound. *See* NUCLEAR STRUCTURE.

When a group of nuclei of a given precursor undergoes its radioactive decay, only a fraction of the product nuclei will emit neutrons because many of the decays may leave the product nuclei in an insufficiently high state of excitation. Hence, there is a neutron yield probability associated with each decay. After neutron emission, the product nucleus may be stable [for example, precursor ^{137}I beta-decays to ^{137}Xe (excited), which emits a neutron going to ^{136}Xe (stable)], but more commonly the product nucleus will undergo further beta decay, and may itself be a precursor for a subsequent delayed-neutron emission.

Table 1 lists known precursors of delayed-neutron emission, together with the respective half-lives $T_{1/2}$ in seconds and the neutron yield probability in percent. Those in the lighter elements have been made by using particle accelerators, and of these, double, as well as single, delayed-neutron emission has been observed following the beta decay of ^{11}Li, ^{30}Na, ^{31}Na, and ^{32}Na. The main groups are in the fission products, but the longest-lived precursor (^{210}Tl) happens to be in natural radioactivity.

The mean energies of the delayed-fission neutrons are lower than those of the prompt neutrons, being usually in the range of a few hundred keV. The actual delayed-neutron spectra usually have several energy groups, depending upon the nuclear energy levels involved in the beta decay of the precursor.

Role in reactor control. In a ^{235}U nuclear reactor, about 0.7% of the fission neutrons are delayed, the others being prompt. In conventional moderated reactors, the prompt neutrons are born, slowed down, and reabsorbed to produce the next fission in a cycling time of 10^{-3} to 10^{-4} s. (In a fast-neutron reactor the time is much shorter.) Consequently, if the reactor were to become overcritical (more neutrons generated per millisecond than are absorbed or leak out), the chain reaction would exponentiate or "run away," and the reactor might quickly overheat and cause a dangerous accident unless the control rods could respond within a few milliseconds to correct the situation. The fortunate presence of the delayed neutrons eases the situation, because so long as the reactor operates within the margin of 0.7% (delayed critical), the control rods can take as long as several seconds to respond; thus the chain reaction comes within the range of easy and leisurely manual or automatic control.

Within the delayed-neutron margin, advantage can be taken of the reactor's behavior to derive a fundamental characteristic of its operation, namely, its reactivity, which expresses the fractional excess over unity of the chain reaction's effective multiplication

TABLE 1. Delayed-neutron precursors

Nuclide	Half-life ($T_{1/2}$), s	Fraction of beta decays that lead to neutron emission, %
Accelerator-produced light elements		
^{8}He	0.12	12
^{9}Li	0.18	35
^{11}Li	0.008	61
^{12}Be	0.011	—
^{13}B	0.017	0.28
^{16}C	0.75	99
^{17}N	4.2	95
^{19}N	0.42	—
^{27}Na	0.30	0.08
^{28}Na	0.031	0.58
^{29}Na	0.043	15
^{30}Na	0.054	33
^{31}Na	0.017	31
^{32}Na	0.014	15
Fission products, light group		
^{79}Ga	2.63	0.098
^{80}Ga	1.66	0.84
^{81}Ga	1.23	12
^{82}Ga	0.60	21.4
^{83}Ga	0.31	43
^{83}Ge	1.9	—
^{84}Ge	1.2	—
^{84}As	5.6	0.61
^{85}As	2.03	53
^{86}As	0.9	12
^{87}As	0.73	44
^{87}Se	5.6	0.19
^{88}Se	1.52	0.6
^{89}Se	0.41	5.0
^{91}Se	0.27	21
^{87}Br	55.6	2.48
^{88}Br	16.0	6.55
^{89}Br	4.4	13
^{90}Br	1.9	22.5
^{91}Br	0.54	16
^{92}Br	0.36	22
^{92}Kr	1.85	0.033
^{93}Kr	1.29	2.0
^{94}Kr	0.21	5.7
^{92}Rb	4.5	0.011
^{93}Rb	5.85	1.3
^{94}Rb	2.76	9.9
^{95}Rb	0.38	8.4
^{96}Rb	0.22	13.3
^{97}Rb	0.17	25.1
^{98}Rb	0.116	15
^{99}Rb	0.066	14
^{100}Rb	0.05	—
^{97}Sr	0.43	0.3
^{98}Sr	0.80	0.36
^{97m}Y	1.13	0.06
^{98}Y	2.1	3.4
^{99}Y	1.4	1.2
Fission products, heavy group		
^{123}Ag	0.39	—
^{127m}In	1.12	< 0.04
^{127}In	3.76	0.68
^{128}In	0.84	0.06
^{129m}In	0.59	0.25
^{129}In	1.26	2.5
^{130}In	0.58	1.40
^{131}In	0.28	1.72
^{132}In	0.16	4
^{133}Sn	1.47	—
^{134}Sn	1.04	17
^{134}Sb	10.4	0.11
^{135}Sb	1.71	15.5
^{136}Sb	0.82	23
^{136}Tl	17.5	0.9
^{137}Tl	2.8	2.5
^{138}Tl	1.4	6.3
^{137}I	24.5	6.5
^{138}I	6.5	5.3
^{139}I	2.31	9.3
^{140}I	0.60	8.8
^{141}I	0.44	26
^{141}Xe	1.73	0.04
^{142}Xe	1.24	0.42
^{143}Xe	0.30	—
^{141}Cs	24.9	0.037
^{142}Cs	1.71	0.093
^{143}Cs	1.78	1.61
^{144}Cs	1.00	3.1
^{145}Cs	0.58	13.6
^{146}Cs	0.335	13.3
^{147}Cs	0.21	25.4
^{147}Ba	0.71	5.2
^{148}Ba	0.47	23.9
^{147}La	4.1	0.5
Natural radioactivity		
^{210}Tl (Ra C″)	78	0.007

constant, k_{eff}; that is, reactivity $= (k_{\text{eff}} - 1)/k_{\text{eff}}$. The reactivity is used by the operators to follow the response of the reactor to physical changes, as in fuel loading, the insertion or extraction of neutron-absorbing samples, or a change in temperature.

If after a period of steady operation (reactivity = 0) a control rod is inserted farther, the reactor's response will consist of two parts: a quick reduction in power associated with the reduction of the population of neutrons undergoing diffusion, with their millisecond times of residence; and a slow response that eventually becomes the half-life of the longest-lived precursor, ^{87}Br. If, on the other hand, the control rod is withdrawn to a new position, the reactor's response is stronger because a positive exponential is involved. There is a brief, transient jump in power due to an increase in the prompt neutron population, but in the delayed-neutron component there is a balance between the decay of existing precursors made at the original neutron flux level and the growth of new precursors made at the new (higher) neutron flux level. This has a controlling effect because the population of delayed-neutron precursors greatly exceeds the population of diffusing neutrons with their millisecond lifetimes. After the transient

TABLE 2. Classification of delayed-neutron emitters into six groups

Group	1	2	3	4	5	6
$(T_{1/2})_i$, s	55	22	6	2	0.6	0.2
λ_i, s^{-1}	0.0126	0.0316	0.115	0.346	1.15	3.46
β_i, yield per 10^4 fissions	5.2	34.6	31.0	62.4	18.2	6.6

passes, the reactor exhibits an exponential rate of increase in power, and the time required for the power to increase by a factor of e (that is, 2.718) is called the stable period, T. This period may be as long as an hour or more, because the balance between precursors in formation and precursors decaying is delicate. The period in turn is related to the reactivity through the inhour equation (inhour = inverse hour, or $hour^{-1}$), which in one form is expressed by the equation below. Here l is the life time

$$\text{Reactivity} = \frac{l/T + \sum_i \left[\beta_i/(1+\lambda_i T)\right]}{1 - \sum_i \left[\beta_i/(1+\lambda_i T)\right]}$$

of diffusing neutrons, and l/T represents the transient effect. The summation represents the delayed-neutron effect, λ_i being the decay constant of the ith precursor [$\lambda_i = 0.693/(T_{1/2})_i$] and β its fractional intensity. (Physically, β_i would include both the fractional decay yields listed in Table 1 and the fission yield of the associated mass chain.) If l and T are expressed in hours, and the λ_i in $hours^{-1}$, the reactivity will be stated in inhours.

Evaluation of the summation is complex if all of the precursors are listed individually, and, in practice, a grouping has been used. The original resolution of the overall delayed-neutron decay curve (1942) yielded four groups, decaying exponentially with half-lives ranging between 55 and 2.5 s, with associated relative intensities. Subsequent work, particularly upon the short-lived emitters, has expanded this to six groups. As an example, six groups that account for the delayed-neutron decay curve in the case of slow-neutron fission of ^{235}U are given in **Table 2**. In this case the total β is 158 delayed neutrons per 10^4 fissions.

To match this grouping, the summation in the inhour formula is taken over six terms, each with its associated λ_i and β_i.

As related to Table 1, group 1 represents almost entirely ^{87}Br, group 2 is mostly a combination of ^{88}Br and ^{137}I, and the other groups each contain several or many precursors having half-lives close to that associated with their group.

As a further simplification (and a rougher approximation), the six groups can be compressed into one. The total β is then used, and the mean λ is an average of the λ_i, each weighted proportionally to $(T_{1/2})_i$. On the other hand, it is also possible to make the summation over the 80 or more individual precursors, and when this is done, reasonable agreement is found with the six groups of the experimentally measured total delayed-neutron decay curve. *See* REACTOR PHYSICS.

Other uses. Delayed neutrons have been used for the detection and assay of fissile materials in several ways. In prospecting for uranium, for example, if a neutron source and a neutron detector are lowered into a borehole, the source can be activated for an interval and then turned off or removed; the detection of delayed neutrons returning to the counter would then show the presence of fissile ores. Alternatively, if an ore sample is exposed to the neutron flux of a reactor and then moved to a neutron detector, its assay of fissile material can be quickly determined; a sensitivity of 1000 delayed-neutron counts per microgram of uranium content has been reported. The content of nuclear fuel in fuel rods can be nondestructively measured by using the delayed neutrons. They are used for localization of cladding failures in reactor fuel elements; when the cladding fails, fission products appear in the cooling stream, and a neutron detector placed downstream, outside the reactor shield, can signal the presence of a cladding failure with much greater certainty than can a gamma-ray detector. (In fact, triangulation methods have been devised whereby the channel that contains the failure can be localized.)

Finally, the delayed neutrons are of some interest in cosmogony, in connection with the r-process. This is a rapid buildup of the elements and isotopes by successive neutron capture and beta decay in an event such as a supernova explosion, in which an immense density of neutrons is temporarily generated. As the nuclides decay toward stability after the event, the presence of delayed-neutron activities can modify the final distribution of relative abundances, and in fact, the inclusion of this refinement into the r-process calculations seems to improve the agreement with the isotope abundances found in nature. *See* NUCLEAR FISSION; NUCLEOSYNTHESIS; RADIOACTIVITY.

Arthur H. Snell

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Deliquescence

The absorption of atmospheric water vapor by a crystalline solid until the crystal eventually dissolves into a saturated solution. This behavior is well known for certain salts such as hydrated calcium chloride, $CaCl_2 \cdot 6H_2O$, and zinc chloride, $ZnCl_2$, but it is a property of all soluble salts in air of sufficiently high humidity.

Themodynamically, the condition for deliquescence is that the partial pressure of the water vapor in the air exceed the vapor pressure (aqueous tension) of the water in the saturated solution of the salt. Then the reaction solid + water vapor → aqueous solution will occur spontaneously. The speed of the process depends upon the rate of diffusion of water vapor into the crystal lattice, crystal size, and other factors. The process will stop when the water vapor in the atmosphere is depleted to the point at which its partial pressure equals that of the saturated solution.

In general, substances which are highly soluble in water have a greater tendency to deliquesce, since concentrated solutions will have a lower vapor pressure. At 77°F (25°C) the vapor pressure of pure water is 23.8 mmHg (3173 pascals), whereas that of a saturated solution of $CaCl_2 \cdot 6H_2O$ is only 7.0 mmHg (933 pascals); this salt then will deliquesce in an atmosphere where the relative humidity exceeds 7.0/23.8 or 30%. For $ZnCl_2$ the situation is much more extreme; it will deliquesce at a relative humidity of 10%. Ordinary sugar (sucrose) at 77°F (25°C) will deliquesce at humidities above 85%.

Crystalline solids also may absorb water by increasing their water of hydration if the dissociation pressure of the hydrated species to be formed is less than the partial pressure of the water vapor. It is this process, not deliquescence, which is the opposite of efflorescence.

Deliquescent substances can be used to remove water vapor from air, although they have no special advantage over substances which merely add water of hydration and remain crystalline. *See* DESICCANT; DRYING; EFFLORESCENCE; SOLUTION; VAPOR PRESSURE.

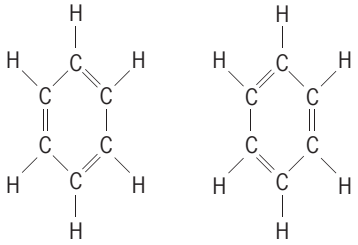
Robert L. Scott

Delocalization

A phenomenon in which the most loosely held bonding electrons of some molecules serve to bind not two but several atoms. This contrasts with localization, a characteristic of ordinary single bonds, as in the normal paraffins, for example, methane (CH_4); in ethane, (C_2H_6); or in the molecules water (H_2O) and ammonia (NH_3). *See* CHEMICAL BONDING.

The prototype completely delocalized system is the ideal crystalline metal, in which the valence electrons are spaced uniformly over a periodic lattice of positive-ion cores. This confers a special stability to the metal, and it also accounts for its high electrical conductivity and other metallic properties. *See* METAL.

The aromatic and conjugated molecules of organic chemistry contain delocalized electronic systems. The benzene molecule (C_6H_6) is considered the archetypical aromatic molecule. Benzene possesses an underlying single-bonded planar framework $(C_6H_6)^{6+}$ plus six additional electrons. Available for these electrons are six $2p$ orbitals, one on every carbon atom, each perpendicular to the molecular plane. The six electrons, called pi electrons, are ascribable to the six carbon atoms in such a way that neither structure (**1**) nor structure (**2**) is an accurate



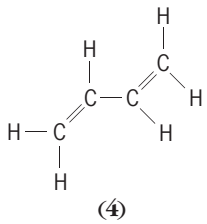
(**1**) (**2**)

representation, but instead a structure that can be described in the valence bond language as a resonance hybrid of the two, schematically written as (**3**), where the circle stands for the delocalized elec-

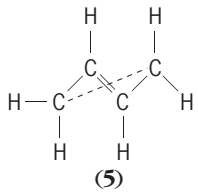
(**3**)

trons. All molecules containing such rings possess an extra stability associated with the delocalization phenomenon. As a consequence, they also have a low propensity for chemical reactivity.

Delocalization is also present, though to a lesser extent, in conjugated molecules such as butadiene (C_4H_6) which contain alternating single and double bonds. Butadiene has the principal localized structure (**4**), but this structure is in resonance with a somewhat less stable structure (**5**). The resonance



(**4**)



(**5**)

gives a certain amount of delocalization. *See* CONJUGATION AND HYPERCONJUGATION; RESONANCE (MOLECULAR STRUCTURE).

The greater the length of a path of electron delocalization, the more the wavelength of light that a molecule can absorb will shift to the red. Thus

butadiene is colorless, but β-carotene is colored; benzene is colorless, but triphenylmethyl dye molecules are colored. Another manifestation of the mobile nature of the pi electrons in aromatic systems is the ring-current effects they produce on proton chemical shifts in nuclear magnetic resonance spectra. *See* MAGNETIC RESONANCE.

When all the bonding in a molecule is localized, it can be stated to good accuracy that many molecular properties (for example, the energy) are sums of bond contributions, and that the properties of individual bonds (for example, polarity and length) can be transferred from one molecule to another. When the bonding is delocalized, on the other hand, simple additivity and transferability rules do not hold.

Quantititative understanding of localization and delocalization has been achieved by using calculational techniques from valence bond and molecular orbital theories of chemical bonding. In the valence bond approach, substantial resonance signifies the presence of delocalization. In the molecular orbital approach, wherever delocalization is large, the highest occupied molecular orbitals will turn out necessarily to extend over several atoms. *See* MOLECULAR ORBITAL THEORY.

Robert G. Parr

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Delta

A deposit of sediment at the mouth of a river or tidal inlet. The name, from the Greek letter Δ, was first used by Herodotus (fifth century B.C.) for the triangular delta of the Nile. It is also used for storm washovers of barrier islands and for sediment accumulations at the mouths of submarine canyons. *See* FLOODPLAIN; NEARSHORE PROCESSES.

Structure and growth. The shape and internal structure of a delta depend on the nature and interaction of two forces: the sediment-carrying stream from a river, tidal inlet, or submarine canyon, and the current and wave action of the water body in which the delta is building. This interaction ranges from complete dominance of the sediment-carrying stream (still-water deltas) to complete dominance of currents and waves, resulting in redistribution of the sediment over a wide area (no deltas). This interaction has a large effect on the shape and structure of the delta body.

Most of the sediment carried into the basin is deposited when the inflowing stream decelerates. If there is little density contrast, this deceleration is sudden and most sediment is deposited near the mouth of the river. If the inflowing water is much lighter than the basin water, for example, fresh water flowing into a colder sea, the outflow spreads at the surface over a large distance away from the outlet. If the inflow is very dense, for instance, cold muddy water in a warm lake, it may form a density flow on or near the bottom, and the principal deposition may occur at great distance from the outlet. A good example is Lake Mead, where 134 ft (41 m) of sediment have been deposited against Hoover Dam at 75 mi (120 km) from the inlet of the Colorado River. However, not all long-distance transport of river sediment in the sea can be attributed to deep density flow; almost half of the total load of the Amazon River is deposited more than 1000 mi (1600 km) away near the Orinoco delta and in the southeast Caribbean due to longshore transport by currents and waves.

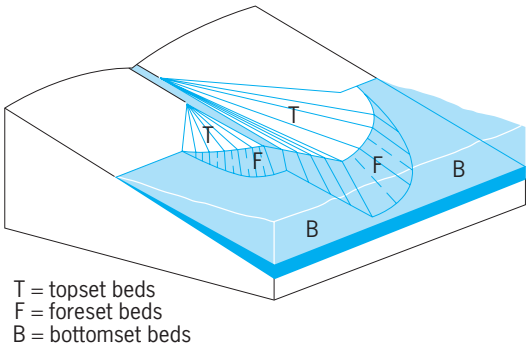

Fig. 1. Schematic diagram showing two stages of growth of a typical Gilbert-type delta. (*After P. H. Kuenen, Marine Geology, John Wiley and Sons, 1950*)

Three principal components make up the bodies of most deltas in varying proportions: topset, foreset, and bottomset beds. They were defined originally by G. K. Gilbert for lake and shallow marine deltas; there, sizable differences in the depositional slopes of the three units, which form part of the definition, are readily seen (**Fig. 1**). The concept has been redefined to apply to most deltas, and as now understood, the topset beds comprise the sediments formed on the subaerial delta: channel deposits, natural levees, floodplains, marshes, and swamp and bay sediments. The foreset beds are those formed in shallow water, mostly as a broad platform fronting the delta shore, and the bottomset beds are the deep-water deposits beyond the deltaic bulge. In marine deltas the fluviatile influence decreases and the marine influence increases from the topset to the bottomset beds. Concurrently, the mean grain size and lithological variability also decrease. The topset beds are variable in grain size and composition over short distances, ranging from coarse channel deposits through layered levee silts and clays to the fine clays of the floodplains and marshes. The foreset beds consist of a uniform blanket of laminated silts and silty clays reflecting the seasonal influence of currents and waves on the shallow platform, while the bottomset beds are homogeneous silty clays, often characterized by abundant plant fragments. The boundary between bottomset beds and neritic shelf clays is completely arbitrary. Similarly, the faunas incorporated in the sediments range from great local variability in the topset beds, with dominance of barren sediments and local pockets of rich brackish faunas, to the almost completely marine fossils uniformly distributed in the bottomset beds.

Through progressive outbuilding, the delta can become overextended with long river courses and very low gradients. Eventually, shorter and steeper paths to the sea will be developed and the existing subdelta will be abandoned in favor of a shorter course. The Mississippi delta shows good examples of such subdelta migration, and the present active delta would be abandoned for a new one off the Atchafalaya River, if artificial control did not keep the flow in check. The abandoned delta body gradually submerges and is eroded by waves and wind. The submergence results from two factors: temporary continuation of subsidence caused by loading of the deltaic sediments on the substrate, and compaction of the fine sediments. These are overcompensated during delta growth, but cause it to sink below sea level after abandonment. Winnowing of the sediments by waves produces a lag deposit of sand at the advancing edge of the sea, which ultimately results in a thin and discontinuous blanket of sand, often bordered by a string of sand bars and small islands near the previous delta shore. This sequence is called the destructive phase of delta formation; a good example is the Breton Sound-Chandeleur Island area of the Mississippi delta, which marks the site of the old St. Bernard subdelta.

In a different way, deltas can be viewed as being composed of three structural elements: (1) a framework of elongate coarse bodies (channels, river-mouth bars, levee deposits), which radiate from the apex to the distributary mouths (sand fingers); (2) a matrix of fine-grained floodplain, marsh, and bay sediments; and (3) a littoral zone, usually of beach and dune sands which result from sorting and longshore transport of river-mouth deposits by waves, currents, tides, and wind. The relative proportions of these components vary widely. The Mississippi delta consists almost entirely of framework and matrix (**Figs. 2** and **3**); its rapid seaward growth is the result of deposition of river-mouth bars and extension of levees, and the areas in between are filled later with matrix. This gives the delta its characteristic bird-foot outline. A different makeup is presented by the Rhone delta (Fig. 3), where the supply of coarse material at the distributary mouths is slow, and dispersal by wave action and longshore drift fairly efficient, so that nearly all material is evenly redistributed as a series of coastal bars and dunes across a large part of the delta front. This delta advances as a broad lobate front, while the present Mississippi delta grows at several localized and sharply defined points.

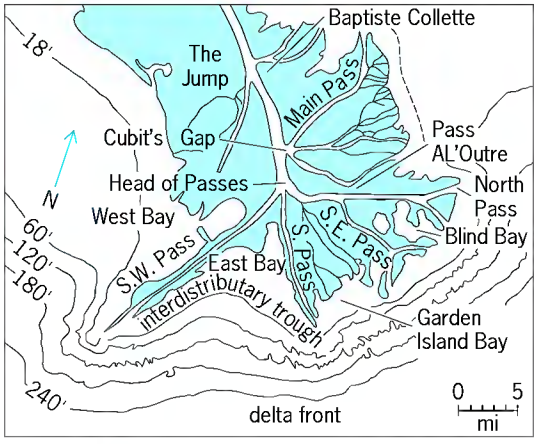

Fig. 2. Modern Mississippi bird-foot delta. 1 mi = 1.6 km. (*After H. N. Fisk et al., Sedimentary framework of the modern Mississippi delta, J. Sediment. Petrol., 24:76–99, 1954*)

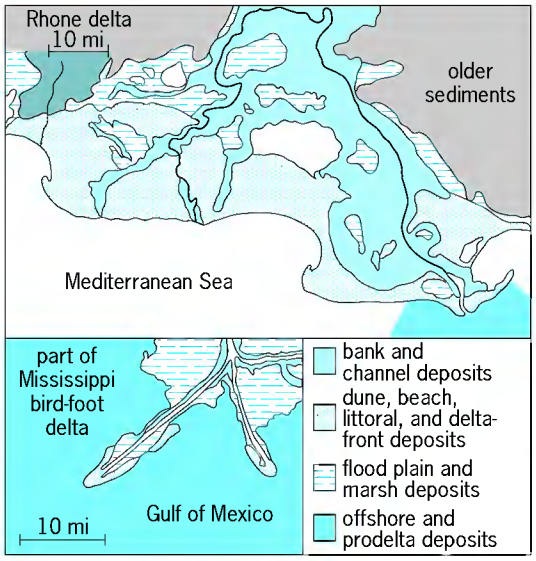

Fig. 3. Examples of deltaic sediment distribution. 1 mi = 1.6 km.

Pattern of deposition. Many different types of deltas exist, few of which have been examined in detail. It appears that most of them can be explained by the interplay, in various proportions, of two major controlling forces: the coarseness and the rate of sediment supply by the river, and the rate of reworking and dispersal along the delta front by marine currents and waves. Other factors, such as the shape of the preexisting coastline and of the sea floor, seasonal fluctuations of marine and river forces, and the tidal range, cannot as yet be evaluated, but appear to be secondary. The Mississippi delta (**Fig. 4*a***) is dominantly fine-grained and consists in large part of thick sand fingers with intervening matrix clays. It builds rapidly outward, is almost entirely controlled by river action, and has a typical bird-foot shape.

In the Rhone delta (Fig. 4*b*) the rate of sediment supply is much lower, but the sediment is coarser, and the rate of reworking and dispersal by waves and currents is higher. Consequently, this delta is more sandy, and its principal growth takes place through the accretion over a broad front of bars and beaches, resulting from winnowing of sand at the river mouth and dispersal by wind and waves. The advance of the delta is slow, but occurs over a broadly lobate front, and the bulk of the delta body consists of the deposits of the sandy strand plain. Matrix and framework deposits are restricted to the topset beds in the upper delta. Littoral marine forces control to a large extent the shape of this delta.

A broad coastal plain is also found in the Orinoco

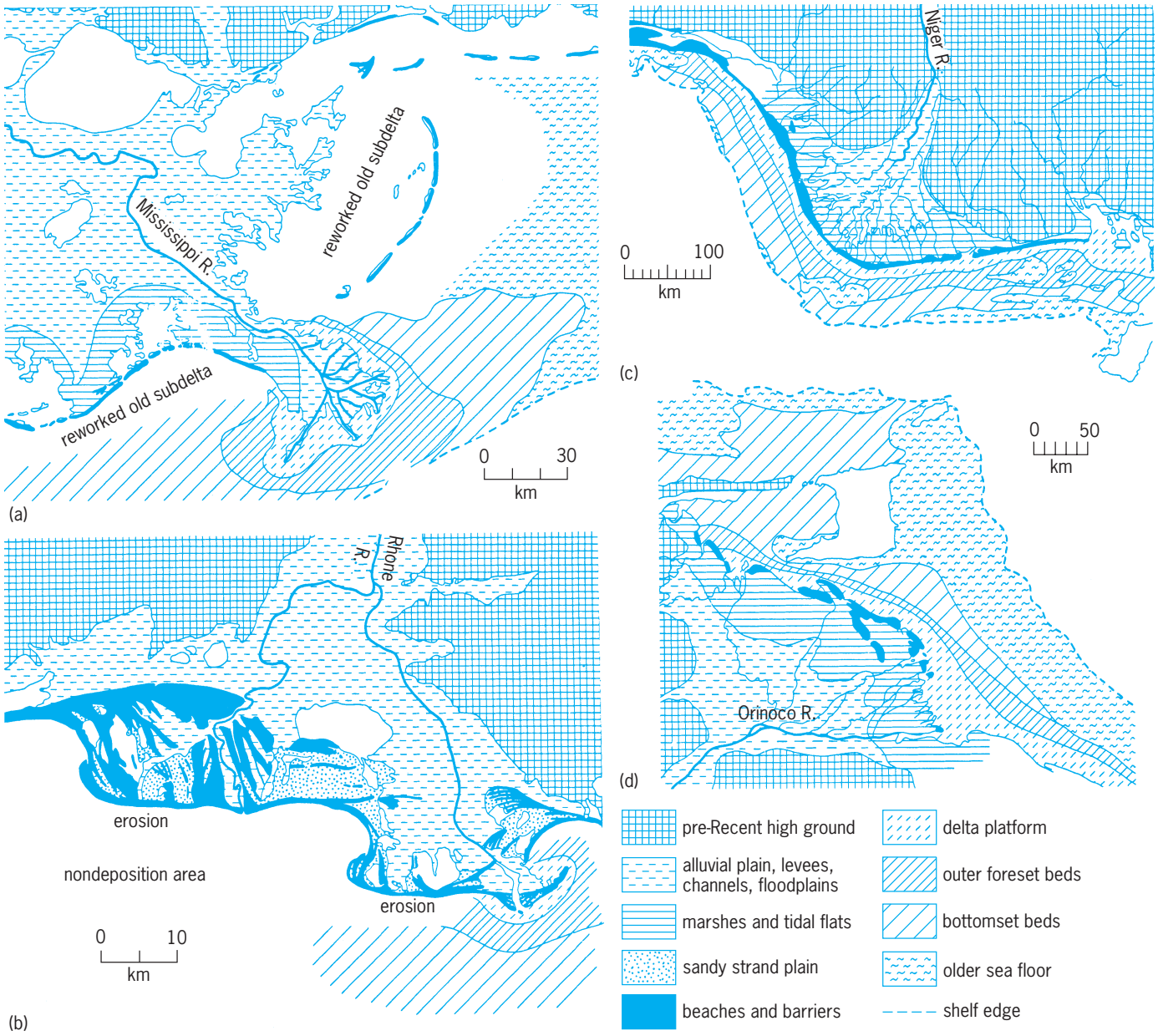

Fig. 4. Sketch maps of the sediments of the (*a*) Mississippi, (*b*) Rhone, (*c*) Niger, and (*d*) Orinoco deltas. 1 km = 0.6 mi. (*After T. H. van Andel, The Orinoco delta, J. Sediment. Petrol., 37(2)297–310, 1967*)

delta (Fig. 4*d*). The large supply of very fine sediment of this river is augmented by Amazon mud brought along the shore from the east. It builds rapidly in a sheltered sea with little wave action, but longshore drift to the northwest is marked. Consequently, the coastal plain, although also littoral marine, is not wave-built and sandy, but formed by tidal deposition on mud flats. The delta, as a result of the longshore currents, advances over its entire perimeter, and the action of the river plays an insignificant part in its development.

Delta advance over the entire perimeter also takes place in the Niger delta (Fig. 4*c*), which resembles the Orinoco delta in shape and distribution of sediments. However, enough sediment is supplied by this river to form, under the influence of moderate wave action on an exposed shelf, a marginal zone of sand bars, beaches, and barrier islands. Behind this barrier, tidal sedimentation of a muddy coastal plain takes place, and framework-matrix units are restricted to the upper delta. The delta advances over a broad front as a result of redistribution of sediments by currents. In a sense it is intermediate between the Orinoco and Rhone deltas.

Characteristics of modern deltas. While over 150 major deltas are formed today, not all rivers, or even all major ones, have deltas. This is the result of a rise in sea level following the last glacial period, which produced deep estuaries in many parts of the world which have not yet been filled (for example, the Amazon estuary). Delta thicknesses vary widely: The Nile is depositing a layer 50 ft (15 m) thick in a

Statistics pertaining to modern deltas*

River	Dimension of subaerial delta, statute mi		Amount of sediment discharged		Annual extension of subaerial delta	
	Length	Breadth	River water by weight (avg), ppm	Annual volume of sediment, mi^3	Measurement period, years	Approximate distance, ft
Mississippi's present bird-foot delta	12	30	550	0.068	1838–1947	250
Hwang-Ho	300	470†	50,000‡		1870–1937	950
Ganges-Brahmaputra	220	200	870	0.043 (Ganges only)		
Rhone into Mediterranean Sea	30	47	400–590	0.005	1737–1870	190
Danube	46	46	310	0.008		40
Nile (prior to barrages)	96	145	1600	0.001	1100–1870	45
Colorado above Hoover Dam	43	0.05–0.6	8300	0.032	1936–1948	3.6 mi (gorge)
Euphrates-Tigris	350	90			1793–1853	180

*1 mi = 1.6 km. 1 mi^3 = 4.17 km^3. 1 ft = 0.3 m.
†Includes 100 mi of nondeltaic Shantung Peninsula.
‡Maximum is 400,000 ppm.

shallow embayment, whereas the Mississippi, building out in deeper water, has constructed a delta body more than 850 ft (260 m) thick. This thick wedge exerts pressure on the underlying beds, causing downbending of the shelf and producing mudlumps in areas where the underlying sediment is soft (see **table**).

Engineering problems. Despite difficult engineering problems, many cities, such as Calcutta, Shanghai, Venice, Alexandria (Egypt), and New Orleans, were constructed on deltas. These problems include shifting and extending shipping channels; lack of firm footing for construction except on levees; steady subsidence, which may reach a rate of 5 ft (1.5 m) per century; poor drainage; and extensive flood danger. Submergence to a depth of 15 ft (4.6 m) during hurricanes or typhoons is not uncommon. Moreover, in certain deltas the tendency of the main flow to shift away to entirely different areas, with resulting disappearance of the main channels utilized for water traffic, is a constant problem that is difficult and costly to counter. *See* ESTUARINE OCEANOGRAPHY; MARINE SEDIMENTS.

Tjeerd H. Van Andel

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Delta electrons

Energetic electrons ejected from atoms in matter by the passage of ionizing particles. In every primary ionizing collision between a charged particle and an atom, one or more electrons are ejected. Delta electrons are, by definition, that small fraction of these emitted electrons having energies which are large compared to the ionization potential. The name is a traditional one—comparable to alpha particles, for energetic helium nuclei, and beta particles, for energetic electrons emitted in radioactive decays. *See* ALPHA PARTICLES; BETA PARTICLES; CHARGED PARTICLE BEAMS; IONIZATION.

The energy W of the delta electron emitted at angle θ in the laboratory is given by Eq. (1), where v is

$$W = 2mv^2 \cos^2 \theta \tag{1}$$

the velocity of the incident charged particle, and m is the electron mass. The electrons in the forward direction are the most energetic and have an energy near $4(m/M)E$, where M and E are the mass and energy of the incident charged particle. The cross section for the ejection of a delta electron in the energy range between W and $W + dW$ given by Eq. (2), where e

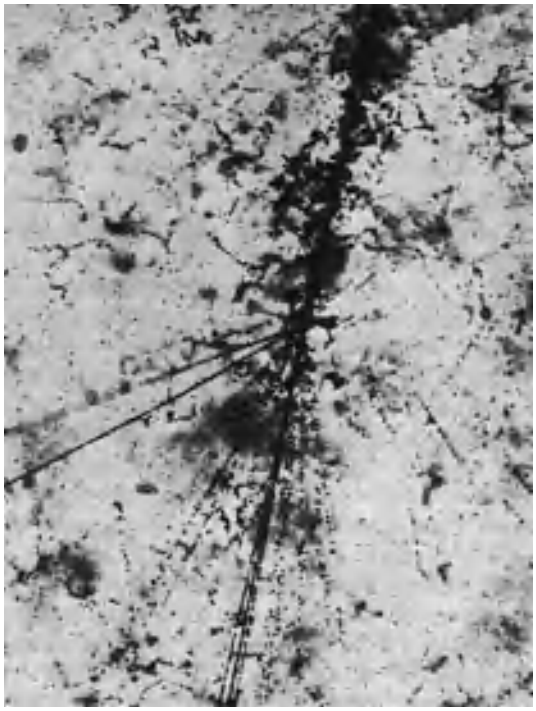
$$d\sigma = \frac{2\pi e^4 z^2}{mv^2} \cdot \frac{dW}{W^2} \tag{2}$$

is the electron charge and ez is the charge of the incident particle, and the cross section for finding an electron between θ and $\theta + d\theta$ is given by Eq. (3).

$$d\sigma = \frac{2\pi e^4 z^2}{m^2 v^4} \cdot \frac{\sin\theta \, d\theta}{\cos^3 \theta} \tag{3}$$

The number of electrons emitted per centimeter of path length is obtained by multiplying these cross sections by the number of electrons per cubic centimeter in the stopping material.

Delta electrons are responsible for the "hairy" appearance of charged particle tracks when they are observed in cloud chambers or in photographic emulsions (see **illus.**). In studies of super-high-energy particles in cosmic radiation and from the highest-energy accelerators, observation of the number of delta electrons per centimeter of path length has been shown to lead to a reliable determination of the charge of the energetic particle. Biological



Nuclear emulsion showing iron nucleus ($Z = 26$) in cosmic radiation entering from upper right and producing large numbers of delta electrons, giving its track a heavy, "hairy" appearance.

response to ionizing radiation is related to the production of delta electrons by the primary ionizing particle. D. Allan Bromley

Delta resonance

A member of a class of subatomic particles called baryons, which exists in four electric charge states and has a total spin of $J = {}^3/_2$. In the underlying quark model, the delta resonance (Δ) consists of three quarks whose intrinsic spins of ${}^1/_2$ are lined up in the same direction. The Δ is closely related to the more familiar nucleon constituents of atomic nuclei, the neutrons (n) and protons (p). *See* NUCLEON; QUARKS.

The Δ was first observed as a resonant interaction of a beam of pi mesons (π) with a proton target. The probability of a scattering interaction between the π and the proton is strongly dependent on energy, attaining a maximum at the Δ mass of 1236 MeV/c^2 (where c is the speed of light). The formation of the very short-lived Δ (with a lifetime on the order of 10^{-23} s) is followed immediately by its decay back into pion and nucleon. *See* MESON; SCATTERING EXPERIMENTS (NUCLEI).

The understanding of some nuclear phenomena requires explicit treatment of π or Δ degrees of freedom. For instance, the Δ modifies the polarizability of the nucleus, and thus contributes to the quenching of the transition strength observed in charge exchange reactions [that is, (p,n) reactions] on nuclei. The Δ also plays a role in the quenching of the beta-decay process ($n \rightarrow p + e^- + \overline{\nu}_e$, where e^- denotes an electron and $\overline{\nu}_e$ an antineutrino) in the nucleus. High-momentum components are induced in the nucleus through the presence of Δ's. These are probed by a variety of transfer processes. The effect of the Δ is also seen in electromagnetically induced reactions [(γ,p), (γ,π^0), (e,$e'p$), and so forth] involving the interaction of a real or virtual photon (γ) with the nucleus. *See* NUCLEAR REACTION; NUCLEAR STRUCTURE; RADIOACTIVITY.

The Δ thus plays an important role in a wide variety of nuclear phenomena, even under conditions of low energy and momentum transfer. The study of these phenomena reveals much about the presence of pions in nuclei, in addition to neutrons and protons. *See* BARYON; ELEMENTARY PARTICLE. Carl Dover

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Demagnetization

The process of reducing the retained level of magnetization (remanence) in an object. Demagnetization is necessary because the ferromagnetic elements iron, nickel, and cobalt and their alloys can become permanently magnetized if they are exposed to a magnetic field. There is a vast range of sizes of objects whose magnetization would degrade their performance. Examples range from mine-sweeping ships, though magnetic shields for cathode-ray tubes, down to even smaller objects. The level retained and the field required to demagnetize depend on their ferromagnetic behavior. Permanent magnets are called "hard" and are required to retain magnetization since it is this property that makes them useful. The magnetic materials used in transformers are called "soft" and here the retention of magnetization could detrimentally alter the performance of the transformer.

Mechanical factors such as residual stress levels also contribute to magnetic behavior. The dimensions of an object are important as are its material properties. If certain permanent magnets are made too thin, they will lose their magnetization. In this case, this unwanted demagnetization is caused by an internal magnetic field whose magnitude is related to the ratio of the cross section of the magnet poles to the thickness. Sometimes when the magnetic field associated with an object must be reduced to an acceptable level and demagnetization in impracticable, coils are added to the object and the current in them is adjusted so as to oppose the field. *See* FERROMAGNETISM; MAGNET; MAGNETIC MATERIALS; MAGNETISM; MAGNETIZATION; TRANSFORMER.

Need for demagnetization. When permanent magnetization remains in a material the object behaves like a magnet and produces a magnetic field in the

surrounding space. In situations where controlled magnetic fields are to be produced, such as magnetic resonance imaging (MRI) scanners, this behavior will be a problem. Also, if the environment in which a magnetometer that is sensitive to very small magnetic fields is housed produces its own magnetic field, then the magnetometer will not work as required. These are examples of situations where, as far as possible, the material selected should not retain magnetization. *See* MAGNETOMETER; MEDICAL IMAGING.

Techniques. Magnetic materials can exhibit an enormous range of behavior, and it is this variety that presents a challenge if they need to be demagnetized. A process that is suitable for one object may not be as effective for another, even if it is only the geometry of the object that has changed.

Use of magnetic fields. When ferromagnetic materials are incorporated in an object, a way of reducing inadvertent magnetization to an acceptable level is required. To demagnetize them, it is necessary to generate a magnetic field that is of sufficient magnitude and can be reduced to a very low value in small steps. The magnitude of the magnetic field should be such that the magnetization of the material is taken as close as possible to saturation. If the applied magnetic field is not strong enough, the success of demagnetizing will be limited. It is necessary to reduce the magnetic field through cycles that take the magnetization through reversals such that a north pole becomes a south pole and vice versa and then back to a north pole. An alternating magnetic field is usually employed and the frequency should be low enough that complete penetration of the magnetic field occurs. If the material is thick (thicker than 20 mm or 0.8 in.) and the electrical conductivity is high, then the frequency used should be low (not greater than 10 Hz). *See* SKIN EFFECT (ELECTRICITY).

A wire carrying a current can produce the magnetic field needed to demagnetize an object. Coils consisting of a number of turns are typical, and the object is placed inside the coil, where the magnetic field is higher. If the object is large, then the coil diameter will need to be large and this reduces the magnetic field at the center of the coil for a given current. If large currents cannot be produced, then the number of turns can be increased. The magnetic field produced will be nonuniform and it may be necessary to use other coaxial coils to improve the uniformity to ensure that the magnetic flux reaches the required value throughout the material volume. If the properties of the material vary with direction, then coils along more than one axis may be necessary.

When reducing the magnetic field it is very important that the change in the amplitude between one cycle and the next is not too large; otherwise, the demagnetization will not be complete. Usually, it is necessary to repeat the demagnetization process a number of times, typically three, and each additional attempt after the first will improve the demagnetization. To determine if the demagnetization is sufficient, the object should be used in the required manner and its performance should be assessed. If unacceptable magnetization remains, then the demagnetization process should be altered with changes made to the frequency and to the rate at which the amplitude of the magnetic field is reduced to zero.

Use of heating. Another way of demagnetizing a material is to heat it above the Curie temperature. When the material is returned to room temperature, the magnetization will have been reduced. When this approach is used, it is essential to assess whether heating to the required temperature irreversibly alters other properties of the material. Some alloys can undergo an internal structure change at a temperature lower than the Curie temperature that would make them unsuitable for the desired application. *See* CURIE TEMPERATURE.

Michael Hall

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Demodulator

A device that recovers from a transmitted signal information that has been embedded in the signal by a process of modulation. The modification of a usually sinusoidal electrical carrier signal (or simply carrier) for the purpose of transmitting information that is carried on a second electrical modulating signal or information signal is called modulation. The device or circuit that performs the modulation is called a modulator. The information may be from any of a variety of sources such as sensor data, speech signals, or image data. The modulator varies the carrier in amplitude, frequency, phase, or some combination to embed the information for efficient transmission. After reception, this information is recovered from the received signal by a demodulator. During modulation, information is shifted in frequency from baseband to the carrier frequency for efficient transmission. Demodulation is the reverse process by which the information is restored in frequency to its baseband. *See* CARRIER; MODULATION; MODULATOR.

One primary class of modulation (and demodulation) is amplitude modulation, which is subdivided into the classes double-sideband transmitted carrier (variously called DSTC, AMTC, or "standard" AM); the more complex but power-efficient double-sideband suppressed carrier (DSSC);

the even more power- and bandwidth-efficient (and complex) single-sideband suppressed carrier (SSBSC, or simply SSB); vestigial sideband (VSB), in which one sideband is only partially removed, leaving only a vestige of itself; and quadrature amplitude modulation (QAM), in which two amplitude-modulated (AM) signals are transmitted simultaneously on orthogonal carriers (that is, 90° out of phase with one another), that consequentially stay out of one another's way while sharing a common spectrum. The other major modulation methods are frequency modulation and phase modulation; both are highly resistant to amplitude noise because the amplitudes of the received signal are forced to a constant value. *See* AMPLITUDE MODULATION; FREQUENCY MODULATION; PHASE MODULATION; SINGLE SIDEBAND.

The information signals usually are of analog origin but may be converted to digital format for more exact processing. Transmission is always analog or continuous, but within the receiver the signal may be converted to digital format by an analog-to-digital converter (ADC) after the IF (intermediate-frequency) amplifier stage because digital signals can be more precisely processed for superior clarity. The post-IF-stage demodulator may be either analog or, if an ADC is used, digital. Most digital demodulation methods mimic their analog counterparts because of the simplicity, reliability, and thoroughly understood behavior. *See* ANALOG-TO-DIGITAL CONVERTER; INTERMEDIATE-FREQUENCY AMPLIFIER; RADIO RECEIVER.

Demodulation of AM-signals. DSTC signals can be recovered by either envelope detection (also known as asynchronous or incoherent detection) or product detection (also known as synchronous or coherent detection). DSTC signals are of the form $\cos(2\pi f_c t) \cdot [A + B \cdot s(t)]$, where f_c is the carrier frequency, t is time, $s(t)$ is the modulating signal with a maximum amplitude of unity, and A and B are positive, real constants, where $A > B$, so that the bracketed quantity is always positive. For that reason, rectification in the recovery process will not introduce signal distortion. Envelope detection is a two-stage process that first rectifies the signal (that is, all negative values are either eliminated or converted to positive values) to produce the information signal, plus the rectified carrier. The rectified carrier appears as noise, which is then removed by lowpass filters. In product detection, the received signal is multiplied by a replica of the carrier (discussed below), thereby generating both the information signal and a double-frequency carrier signal, which must be removed by a lowpass filter. All AM signals, including single-sideband signals, can be demodulated by product detection followed by a lowpass filter. *See* AMPLITUDE-MODULATION DETECTOR; ELECTRIC FILTER; RECTIFIER.

Demodulation of QAM signals. A QAM transmitter modulates the analog or digital signal in both amplitude and phase. For digital signals, a so-called M-ary (that is, each data symbol can be represented by one of M possible states in the complex plane) digital QAM signal occupies the same bandwidth as a phase-shift keying (PSK) signal (discussed below), but the power efficiency is greater. The two carrier-replica demodulation signals are orthogonal or in quadrature (that is, 90° out of phase with one another). An M-ary digital QAM signal is generated by segmenting the input-signal data into M-bit lengths. These are divided into two sets, A and B, of $M/2$ bits each. A common segmentation approach is to steer the odd-numbered bits of the input-signal data into the A set and the even-numbered bits into the B set. This pair of $M/2$-bit words is modified by complementing the most significant bit of each word and appending a "1" at the end to provide a pair of $M/2 + 1$ bit words which will be treated as 2's complement signed numbers, passed through a pair of $M/2 + 1$ bit digital-to-analog converters (DACs), and then multiplied (modulated) by $\cos(2\pi f_c t)$ for the A set and by $\sin(2\pi f_c t)$ for the B set. These two sets of analog information are then summed for transmission. *See* DIGITAL-TO-ANALOG CONVERTER.

This seemingly arcane exercise provides a rectangular M-by-M array of evenly spaced points in the complex plane, centered at the origin, that represent the M^2 possible data states. Of course, $M = 2$ is binary. Digital QAM with $M = 4$ is robust and commonly used. As M increases, both the channel capacity and the sensitivity to noise similarly increase.

A QAM digital demodulator, or detector, synchronously demodulates the received signal in two parallel signal paths with the sine and cosine carrier replicas, integrates them for $M/2$ clock periods, samples the signals at the end of the integration intervals, passes the signals through a pair of ADCs which then strip off the least-significant bits and complement the most-significant bits, and interleaves the two resulting A and B bit streams to reconstruct the original digital data.

Phase-locked loops. All synchronous demodulation or detection methods depend upon the availability of a replica of the carrier. Exact frequency is demanded. Small phase errors which cause signal loss can be tolerated except in the case of QAM, where phase inaccuracies cause signals from each channel to bleed into the other. The most powerful and successful device to provide the necessary reference signals is the ubiquitous phase-locked loop (PLL).

Briefly, a PLL is a closed-loop system consisting of a voltage-controlled or numerically controlled oscillator (VCO or NCO) whose output is the sought-after frequency reference for synchronous demodulation. The oscillator output is combined with the signal to be demodulated in a phase detector. The phase-detector output, which is proportional to the phase difference between the two signals, is applied to a loop filter, which provides servo stability for the loop and integrates the signal to provide the command signal for the controlled oscillator. Depending on the sophistication of the design, the PLL can be as perfect in performance as the designer wishes. The gain of the loop filter determines the speed of response of the PLL. For amplitude demodulation, a

slow loop that clings tightly to the carrier frequency and ignores all modulation information is desirable. *See* PHASE-LOCKED LOOPS.

Demodulation of FM signals. Frequency-modulated (FM) signals are in the form below, where the constant k is the modulation index. A useful tool is based on the observation that the time derivative of cos (xt) is $x \cdot \sin(xt)$, which is just the 90° phase-shifted signal scaled in amplitude by the frequency. Consequently, one form of FM demodulation, or discrimination (because this process involves discriminating between frequencies), is time differentiation followed by amplitude detection. A suitable approximation to time differentiation over a narrow frequency band of interest is a linear filter whose amplitude response increases linearly with frequency. A second form of FM demodulation uses a wide-bandwidth PLL which is able to follow the FM signal. The PLL oscillator control input signal is the recovered frequency-demodulated signal, $s(t)$.

$$\cos\left\{2\pi \cdot \int_0^t [f_c + k \cdot s(\tau)]\, d\tau\right\}$$

Transmission of digital data by frequency modulation is called frequency-shift keying (FSK). The frequency of the transmitted signal is determined by the digitized information signal. M-ary FSK employs M distinct frequencies. Binary FSK demodulation consists of determining if frequency A or frequency B is present. The simplest noncoherent approach (no carrier reconstruction is required) applies the signal to a pair of band-pass filters, one tuned to A and the other to B; each filter is followed by an energy detector, such as the cascade of a rectifier, integrator, and threshold. The greater output, A or B, is declared present. M-ary FSK demodulation is an obvious extension. *See* FREQUENCY-MODULATION DETECTOR.

Demodulation of PM signals. Phase-modulated (PM)—sometimes called phase-angle modulated—signals are in the form $\cos[2\pi f_c t + ks(t)]$. Since phase is the time integral of frequency, the demodulated PM signal can be recovered simply by integrating the output of an FM demodulator. Alternatively, the output of the PLL phase detector is proportional to the sine of $s(t)$. The PLL phase detector can be made into a linear phase demodulator by converting its inputs from sinusoids to square waves, but at the price of introducing a small amount of phase jitter in the PLL performance.

Transmission of digital data by phase-angle modulation is called phase-shift keying (PSK), and employs a single frequency. The phase of the transmitted signal is determined by the digitized information signal. Each data symbol in an M-ary PSK system may be represented by a point on a unit circle because the amplitude is held constant. The points are separated in angle (phase) by $360°/M$. In general, PSK demodulation is more difficult than FSK demodulation because of synchronization problems, but changes in state can be detected simply by delaying the signal by one clock period and comparing the delayed and undelayed signals in a correlator. A high output means that no transition has occurred. *See* PHASE-MODULATION DETECTOR.

Modems. Integrated digital communications sets are referred to as modems, for modulator/demodulator. Data modems are highly specialized devices. Almost all data modems today are adaptive, which means that they automatically adapt to variations in the transmission medium. *See* MODEM.

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Demospongiae

A class of the phylum Porifera, including sponges with a skeleton of one- to four-rayed siliceous spicules or of spongin fibers or both. Several genera lack a skeleton, and it is through a study of these seemingly primitive forms that the complicated structure of most adult Demospongiae may be understood. The Demospongiae constitute the most abundant and widely distributed group of sponges, occurring in the sea from the tidal zone down to abyssal depths (at least to 18,000 ft or 5500 m). Three families have invaded fresh water. The species vary in size from thin encrustations several centimeters in diameter to huge cake-shaped forms which may measure up to as much as 6.6 ft (2 m) in diameter.

Comparative studies of the embryology and early attached stages of sponges of the class Demospongiae suggest at least two evolutionary lines within this group. One line of development, the subclass Ceractinomorpha, is characterized by the presence of incubated parenchymella larvae; the other, the subclass Tetractinomorpha, includes oviparous species with parenchymella larvae. However, the more primitive families have incubated amphiblastulalike larvae.

Ceractinomorpha. Among the Ceractinomorpha, the genus *Halisarca*, lacking skeletal elements, is a primitive form. The larva of *Halisarca* is a diploblastula or parenchymella with an outer layer of flagellated cells and an inner mass of presumptive ectomesenchymal cells (**Fig. 1***a*). The out flagellated cells lose their flagella, migrate into the interior (Fig. 1*b* and *c*), and later differentiate into choanocytes. Other cell types characteristic of the adult sponge differentiate, and inhalant canals begin to form. The young sponge soon develops a single internal cavity lined with choanocytes (Fig. 1*d*), and an oscular opening breaks through at the apex (Fig. 1*e*). At

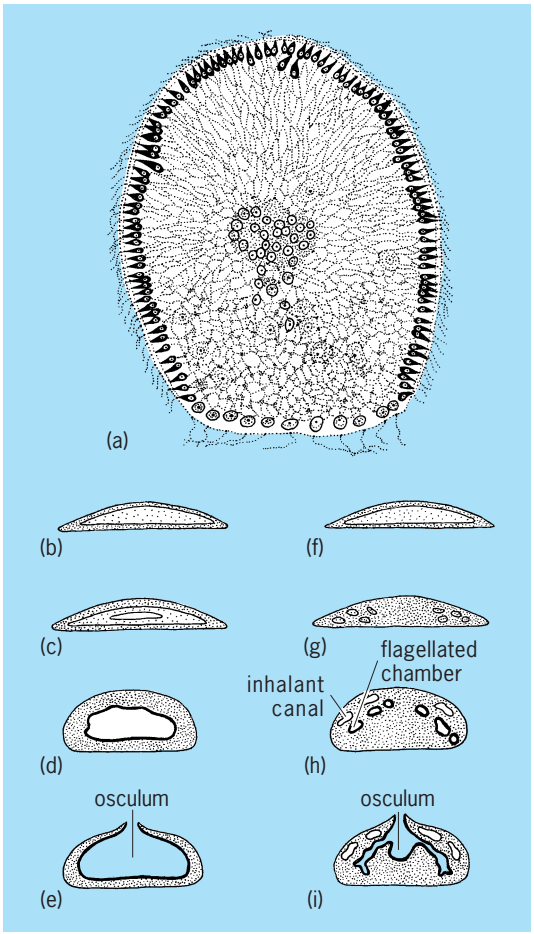

Fig. 1. Development of ceractinomorph sponges. (*a*) Free-swimming parenchymella larva of *Halisarca*. (*b–e*) Metamorphosis of *Halisarca*: (*b*) newly settled parenchymella; external flagellated cells migrate internally; (*c*) internal cavity appears; (*d*) choanocytes line the central cavity; and (*e*) osculum breaks through and young asconoid stage is formed. (*f–i*) Metamorphosis of *Aplysilla*: (*f*) newly settled parenchymella; external flagellated cells migrate internally; (*g*) islands of choanocytes form in the internal mass of cells; (*h*) flagellated chambers and inhalant canals appear; and (*i*) flagellated chambers join exhalant canal system which opens through osculum; young syconoid sponge is formed.

this stage of development, called the rhagon stage, the young *Halisarca* is essentially identical with the asconoid grade of construction found in some Calcarea. Later, folds in the choanocytal layer lead to the formation of flagellated chambers, and a transitory syconoid grade of construction exists. Eventually the flagellated chambers are isolated from one another through the appearance of exhalant canals which converge on the oscula. The adult sponge has a simple leuconoid structure with elongate flagellated chambers having wide openings into exhalant canals and communicating with the surface pores by means of an inhalant canal system. *Aplysilla*, a closely related genus of sponges with a branching, nonanastomosing fibrous skeleton, has a similar developmental history (Fig. 1*f–i*), except that the earliest rhagon stage is syconoid in structure with a folded choanocytal layer. *See* CALCAREA.

Metamorphosis from the larval condition in the Demospongiae characteristically involves a transitory rhagon stage with a simple leuconoid canal system. During further development, the flagellated chambers are isolated between the inhalant and exhalant canals to produce leuconoid canal systems of varying complexity (**Fig. 2**).

Ultrastructural studies of the choanocytes and choanocyte chambers of representative genera of the Demospongiae have revealed differences among the orders of the class. In some genera of the orders Haplosclerida and Petrosiida, belonging to the subclass Ceractinomorpha, the choanocyte chambers lie across the inner ends of the inhalant canals (**Fig. 3**) and are separated therefrom by a pinacocyte layer that is perforated beneath the prosopyles of the

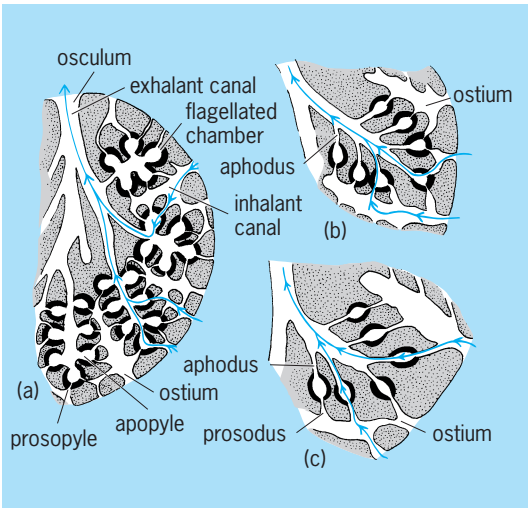

Fig. 2. Types of leuconoid canal systems. (*a*) Eurypylous chambers opening directly into exhalant canal. (*b*) Aphodal chambers; a narrow canal or aphodus leads from the chamber to exhalant canal. (*c*) Diplodal chambers; narrow canal or prosodus intervenes between inhalant canal and chamber, as well as between chamber and exhalant canal. Arrows show direction of water flow.

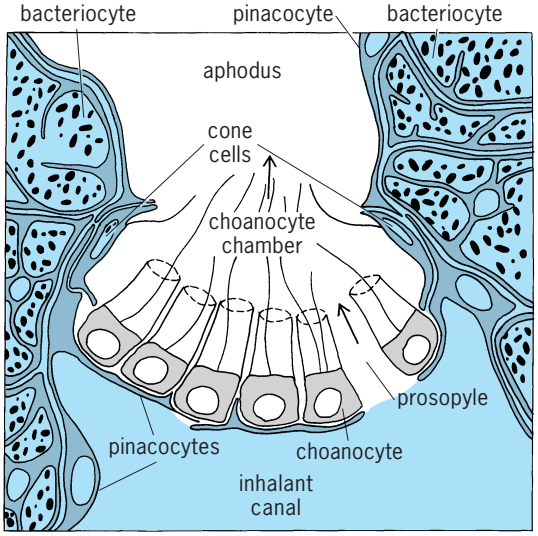

Fig. 3. Section of choanocyte chamber of *Petrosia ficiformis*, showing pathway of water flow from end of inhalant canal through chamber to aphodus.

chamber to allow water to pass from the inhalant canal into the chamber. Between the chamber and the exhalant canal lies a pore that allows water to escape from the choanocyte chamber into an aphodus (Fig. 2*b*) of the exhalant canal system. This pore is surrounded by a ring of cells (called cone cells from their shape) that are part of the choanocyte chamber. This quite complicated structure is, so far as is known at present, restricted to demosponges of the orders mentioned above.

In form, ceractinomorph sponges vary from encrustations, either thin or massive, to lobate and upright branching colonies. The shallow-water species tend to be more plastic in form than the deep-water species, which usually exhibit little intraspecific variation in shape.

Tetractinomorpha. In the other subclass of Demospongiae, the Tetractinomorpha, the skeletonless genus *Oscarella* has a primitive structure. Cleavage results in a solid mass of cells (morula) which later becomes hollow by cytolysis of the interior cells, rich in food reserves. Upon being freed from the parent, the hollow larva is made up of a single layer of flagellated cells and is known as an amphiblastula (**Fig. 4***a*). After a short free-swimming period, the larva attaches to the substrate by its anterior pole and flattens out (Fig. 4*b*). Gastrulation occurs at this point as the anterior half of the larva invaginates (Fig. 4*c*). The blastopore closes as the edges of the now double-layered larva push together against the substratum (Fig. 4*d*). The internal layer of cells is now thrown into folds which pinch off to form cavities which will become the flagellated chambers. Simultaneously a depression forms in the apical ectoderm, and the rudiment of the exhalant canal system appears (Fig. 4*e*). The depression deepens to form a cavity giving off branches which push their way among the flagellated chambers (Fig. 4*f*). The latter finally open into these cavities which become the exhalant canals; inhalant canals push in from the surface of the sponge and join the flagellated chambers. The metamorphosed larva of *Oscarella* assumes the leuconoid grade of construction from the start, with isolated flagellated chambers communicating with inhalant and exhalant canals. The adult *Oscarella* and a related genus, *Plakina*, with two-, three-, and four-rayed spicules, retain the simple leuconoid structure. The sponge consists of a folded wall, each fold made up of a dermal layer and a group of flagellated chambers opening into an exhalant canal. In a more complicated stage, as seen in *Plakortis*, the dermal membrane spreads over the outer ends of the folds, and subdermal cavities are developed. In many species of Tetractinomorpha an extensive cortex is developed, consisting of a network of fiber cells, called desmacytes, which in some cases is overlaid by a thick gelatinous layer containing amebocytes. In form, the species of these orders may be thinly encrusting or massive, but often they have spherical or ovoid shapes. Branching species rarely occur.

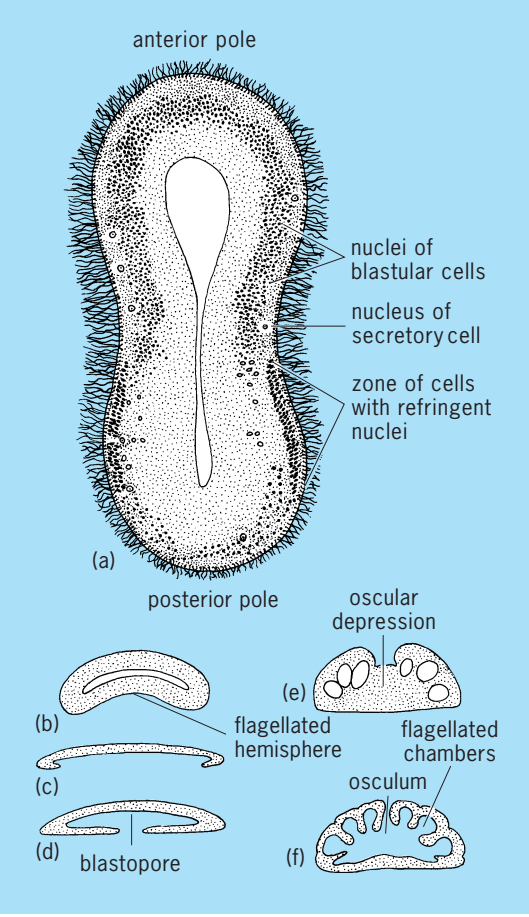

Fig. 4. ***Oscarella***. **(*a*) Free-swimming amphiblastula larva. (*b*) Newly settled larva. (c) Gastrulation begins by invagination of flagellated hemisphere. (*d*) Later stage of gastrulation; blastopore closing. (e) Formation of osculum. (*f*) Young leuconoid stage is formed.**

Skeleton. The skeletal system consists of spicules (sclerites), spongin fibers, or both.

Spicules. The spicules of Demospangiae are intracellular secretions of scleroblasts, cells derived from archeocytes. Each monaxonid spicule is laid down within a single cell by continuous deposition of hydrated silicon dioxide around an organic axial fiber. Above a minimal concentration of silica in the surrounding medium, the length of spicules is independent of the amount of silica present. The thickness of the spicules varies up to a maximal value in correlation with silica concentration, however. Tetraxonid spicules are apparently also formed in individual scleroblasts. Microscleres are formed in special scleroblasts; they require a higher minimal content of silica for formation than do megascleres (**Fig. 5***a–d*).

Fibrous skeletal elements. The reticulum of prominent spongin fibers that forms the skeleton of members of the order Dictyoceratida (bath sponges and related forms) is composed of a felt-work of collagen fibrils, without periodic striation, laid down by specialized cells, the spongioblasts (Fig. 4*e*). Similar spongin fibers enclose or join together siliceous spicules in sponges of the orders Haplosclerida and Poecilosclerida. Intercellular collagen fibrils occur in Demospongiae as well. These fibrils, laid down by collencytes, lophocytes, or pinacocytes, have a

periodic striation and are oriented in a parallel fashion in loose fascicles.

Ceractinomorph skeletal elements. Among the ceractinomorph sponges, spongin tends to be of common occurrence. In the orders Dendroceratida and Dictyoceratida, the latter of which includes the commercially valuable bath sponges, spicules are absent, and the skeleton is formed of spongin fibers only. In the orders Haplosclerida and Poecilosclerida, varying quantities of spongin occur along with siliceous monaxonid spicules. In some species a network of spongin fibers occurs in which the spicules are embedded; in others spongin serves as an interspicular cement. In the order Halichondrida, spongin is rarer in occurrence but is always present in the form of short tracts or as a cement. The cement helps to unite the irregularly arranged siliceous monaxonid spicules.

Ceractinomorph microscleres are commonly C- or S-shaped (sigmas), bow-shaped (toxas), or anchor-shaped (chelas), or are fine and hairlike (raphides).

Tetractinomorph skeletal elements. Tetractinomorph sponges of the orders Homosclerophorida, Choristida, and Spirophorida have little or no spongin and almost always have some tetraxonid siliceous megascleres to which may be added monaxonid types. When triaenes are present, they usually occur in tracts radiating from the central part of the sponge to the surface.

The orders Hadromerida and Axinellida are characterized by a skeleton of monaxonid siliceous megascleres accompanied by varying quantities of spongin. The spicules characteristically occur in tracts radiating from the central regions of the sponge, or in the case of those with an abundance of spongin a plumose arrangement is found. Microsclere types include asters, with numerous rays diverging from a central point; streptasters, spiny rods, often twisted spirally; sigmas; and raphides. The same range of shapes occurs as in other tetractinomorph orders, and in addition many species have an upright branching form. *See* HADROMERIDA.

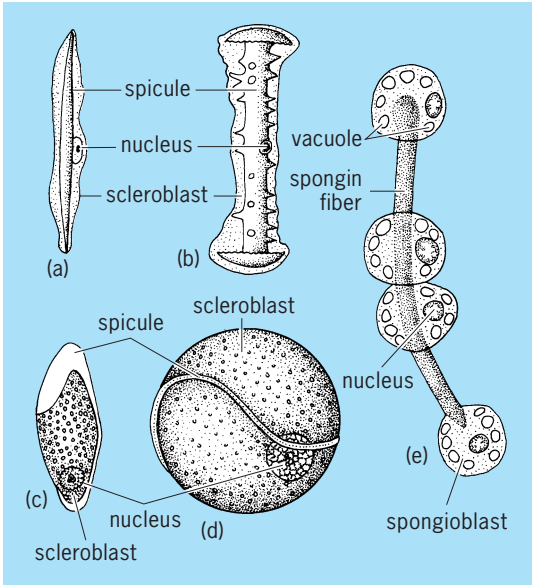

Fig. 5. Skeleton formation in Demospongiae. (*a*) Diactinal megasclere. (*b*) Amphidisc forming within a scleroblast. (*c*) Chela formation within scleroblasts. (*d*) Sigma formation within scleroblasts. (*e*) Spongin fiber in process of formation by spongioblasts in *Haliclona*.

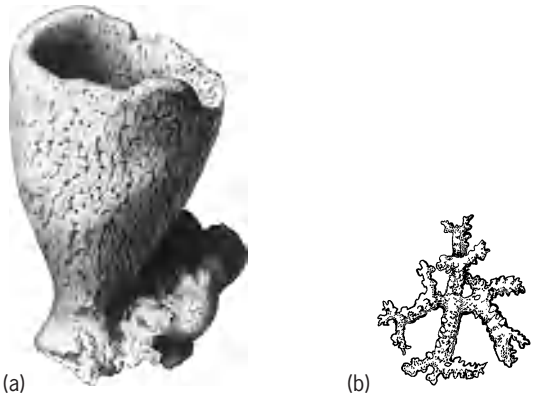

Fig. 6. *Discodermia ornata*, a lithistid sponge. (*a*) Whole sponge. (*b*) Desma.

Lithistid sponges. Some species of Demospongiae are characterized by the presence of spicules called desmas (**Fig. 6**). These are formed by the secondary deposition of silica around ordinary monaxonid or tetraxonid spicules. Supplementary knobby branches often develop, and articulating processes may occur by which neighboring desmas become interlocked to form a stony or lithistid skeleton. Because of the rigidity of the skeletons of lithistid sponges, they are commonly preserved as fossils and are the best-known Demospongiae in the fossil record. Paleontologists have tended to classify such sponges in an order Lithistida. It is apparent from studies of Recent species with a lithistid skeleton that this modification has arisen many times in the evolution of the Demospongiae and that the order Lithistida has no validity. It is difficult to place fossil lithistids among the several orders of the class Demospongiae, however, unless developmental stages of the peculiarly modified spicules are present.

Sclerosponges. While some authors regard the sclerosponges as a subclass of the Demospongiae, they are here regarded as a separate class.

Willard D. Hartman

Dendritic macromolecule

A large molecule having a well-defined three-dimensional structure. Dendritic macromolecules play a crucial role in the chemistry of living systems. In contrast to the high level of structural precision that characterizes many biologically active macromolecules, the sizes and shapes of macromolecules made by polymer chemists are usually far less controlled. Most synthetic polymers are best described as statistical mixtures. However, chemists have sought to develop ways to prepare large molecules with more control over their architecture. If properly designed, such molecules might be capable of performing chemical or physical functions

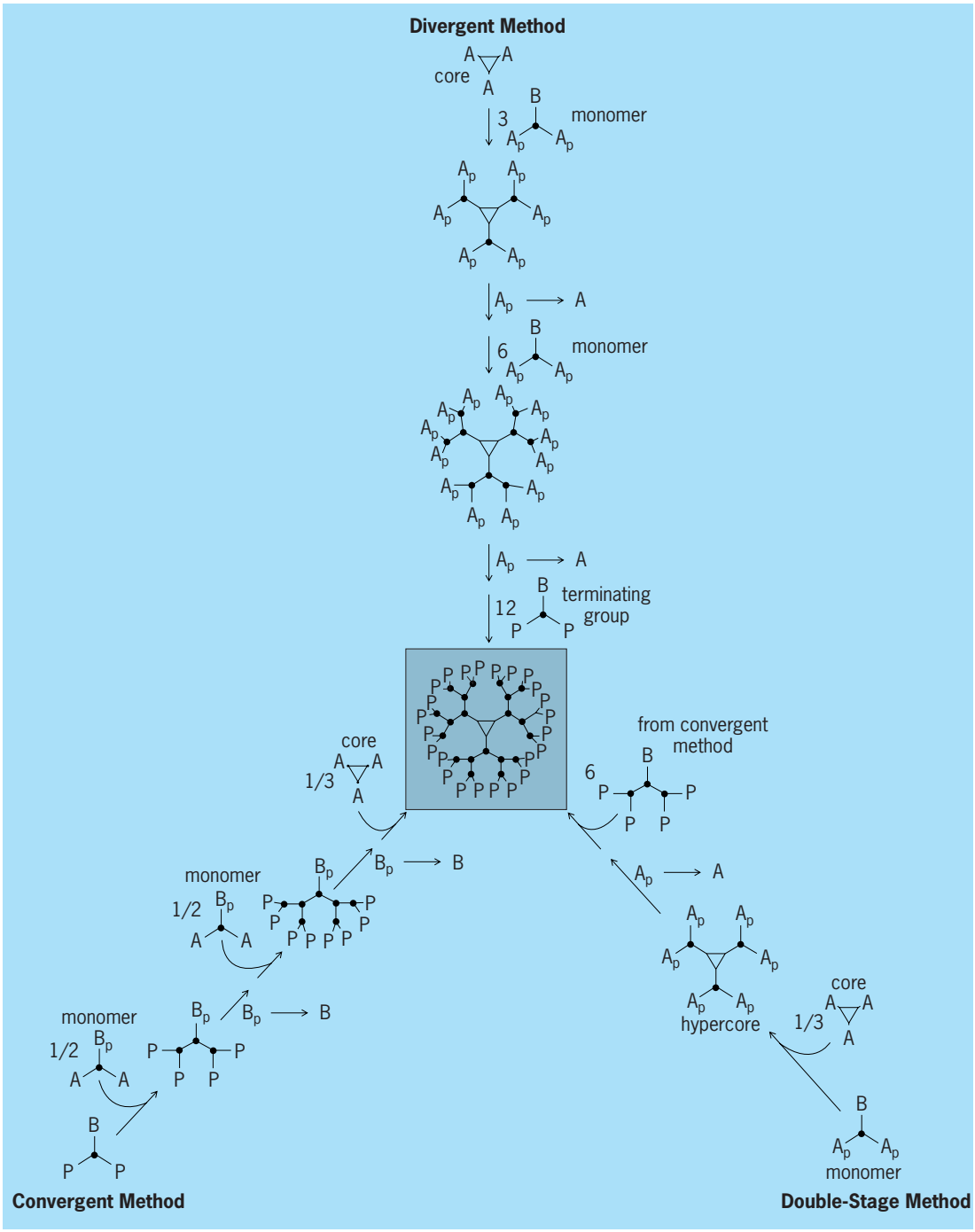

Fig. 1. Diagram of the three synthetic methods used to prepare structure-controlled dendrimers with triconnected branch junctures. Symbols are explained in the text.

reminiscent of the macromolecules found in living systems.

Terminology. Dendritic macromolecules are characterized by a highly branched molecular connectivity, whereby each repeat unit (•) forms a branch juncture (**Fig. 1**). The monomers used to prepare dendrimers possess three or more functional groups, and are of the type AB_2, AB_3, and so forth, where A and B represent a functionality (a site of chemical activity) that can combine to form a new covalent bond. Monomer chemistry is thus similar to that used to make condensation polymers except that the functionality is higher (there are more sites of chemical activity). Repeat units in dendritic macromolecules can be viewed as being arranged in concentric shells centered on a core (▽). By analogy to a genealogical tree, each shell represents a generation of repeat units. The generation farthest from the core makes up the peripheral functionality (P). Dendrimers that emanate from a core of single, double, triple, etc., connectivity are referred to as monodi-, tri-, etc., dendrons, respectively. Monodendrons

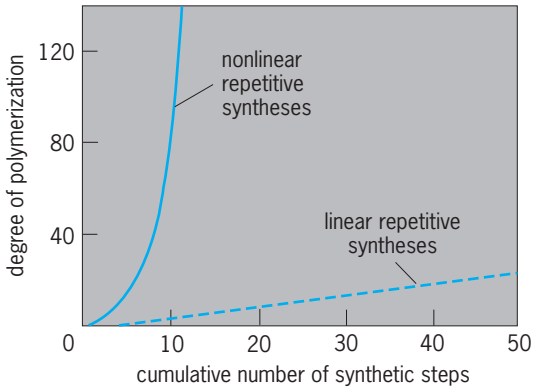

Fig. 2. Plot of the degree of polymerization versus the number of synthetic steps for a typical dendrimer synthesis: nonlinear repetitive synthesis. The efficiency of the dendrimer synthesis is compared to a typical linear repetitive synthesis in which only a constant number of monomers is added during each generation.

have also been called dendritic wedges, and the monodendron core is known as the wedge focal point. *See* POLYMER.

Synthetic strategies. There are two distinct categories of synthetic processes lead to dendritic macromolecules. In the first category, uncontrolled self-condensation of AB_n ($n \geq 2$) monomers generates dendriticlike macromolecules (the one-step method). In the second category, protecting-group chemistry is used to harness monomer condensation so that macromolecules with potentially perfectly defined composition and constitution can be produced (the stepwise, repetitive method). In the stepwise process, monomers are added one generation at a time. The number of monomers added per generation increases in a nonlinear fashion, making the synthesis of these large molecules not only feasible but highly efficient (**Fig. 2**). This type of growth process can be contrasted with a linear repetitive synthesis, in which a constant number of monomer units is added in each round of the synthesis.

Many different dendritic systems have been made by the one-step approach, including several systems that have counterparts as engineering resins. The list includes polyesters, polyamides, poly(siloxysilanes), polyethers, and poly(ether ketones). Like traditional synthetic macromolecules, the size and shape of dendrimers made by the one-step method are best described as statistical mixtures. These dendrimers can be viewed as structurally irregular dendritic wedges with multiple reactive B groups on their periphery and a single reactive A group at their focal point. In these systems an important chemical defect that is thought to limit molecular weight involves intramolecular cyclization of the focal-point functional-group backbiting (an intramolecular chemical reaction) onto one of the many peripheral B groups.

Divergent method. Stepwise synthesis of structurally well-defined dendritic macromolecules typifies the divergent method of dendrimer preparation (Fig. 1). Growth begins at the core and proceeds outward. Because all of the A functional groups on the monomer are protected and therefore not reactive (that is, A_p), growth stops once all of the unprotected factional groups have been consumed. The next round of synthesis can begin only after the A_p groups on the dendrimer's periphery are transformed to their reactive counterparts (that is, $A_p \rightarrow A$). Many different types of dendritic macromolecules have been prepared by the divergent method, including polyamides, poly(amido alcohols), poly(amido amines), polyethers, poly(siloxanes), poly(carbosilanes), and poly(alkanes).

Convergent method. A second route for preparing structure-controlled dendrimers by the stepwise process is the convergent method (Fig. 1). In this method, growth begins from the periphery and proceeds inward, first yielding a series of monodendrons. At the end of the synthesis, the focal point of these monodendrons can be coupled around a core of desired functionality to provide the final dendrimer. The advantage of the convergent method is that the number of reactive groups is constant throughout the synthesis while the molecular weight doubles at each stage. A disadvantage is that the focal point tends to become sterically encumbered, making the synthesis of higher generations more difficult. Dendritic systems prepared by the convergent approach include polyethers, polyesters, polyamides, poly(phenylenes), and poly(phenylacetylenes).

To overcome the problem of steric crowdedness associated with the convergent method, a hybrid approach known as the double-stage convergent method has been described (Fig. 1). In this method, smaller dendritic wedges prepared by the normal convergent approach are coupled to a multifunctional core (hypercore). The hypercore itself can be prepared by either a divergent or convergent process.

An alternative convergent method that alleviates the problem of steric crowdedness involves the use of monomers of progressively larger size at each generation of dendrimer growth. One advantage of this method is that dendrimers having very large dimensions can be assembled rapidly, as has been demonstrated by the ten-step synthesis of a phenyl-acetylene dendrimer that spans more than 12.5 nanometers.

Architecture. The architectures of dendritic macromolecules are dramatically different from those of conventional macromolecules. Variations in molecular size, shape, and flexibility are all possible, depending on the monomer's chemical structure and geometry, the branch-point multiplicity, and the number of repetitive cycles carried out. The architecture most typical of dendritic macromolecules is a globular shape where segment density increases in going from the core to the periphery. The unusual structure of dendritic macromolecules, yet to be fully investigated, results in rheological and solubility characteristics that are dramatically different from linear macromolecules. *See* MACROCYCLIC COMPOUND; ORGANIC SYNTHESIS.

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Dendrochronology

The science that uses annual tree rings dated to their exact year of formation for dating historical and environmental events and processes. Trees are intimately bound to the environment, making them one of the most consistent and dependable recorders of processes and events that occur in nature. Trees, like most plants, are sensitive to both natural (precipitation and temperature patterns) and human-related (air and water pollution) events that trigger certain responses in the vigor of the tree as seen in its growth rate. In most geographic regions, climate patterns in any year cause a response by trees in the volume of wood the tree produces, and often leave indelible "fingerprints" in certain physical and chemical properties of the wood. These fingerprints can be seen in the varying widths of tree rings. In some years, environmental conditions may be favorable for tree growth, allowing trees to produce greater volumes of wood. In other years, climate conditions may be generally unfavorable for tree growth, causing a reduction in the volume of wood produced. *See* DENDROLOGY; TREE; TREE GROWTH.

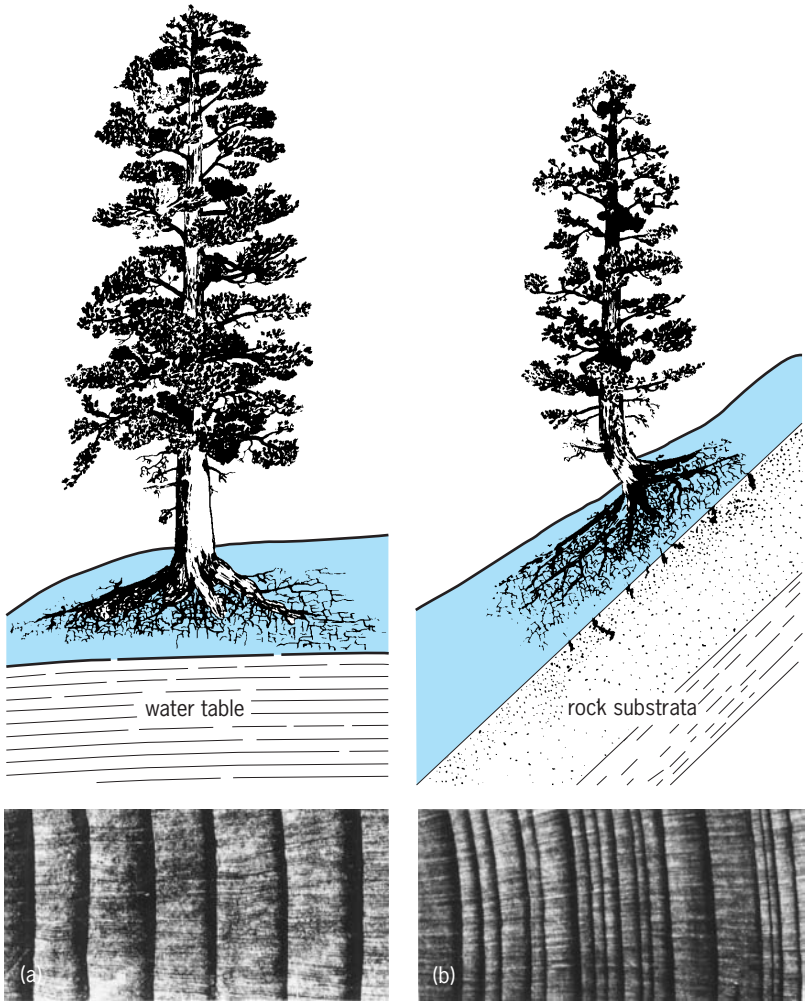

Fig. 1. Ring patterns showing complacent and sensitive ring series. (*a*) Climate has minimal effect on ring widths when trees have ample moisture. (*b*) Maximum variations occur on sites where factors favorable to growth are limited. (*Laboratory of Tree-Ring Research, University of Arizona*)

History. In the early 1900s, Andrew E. Douglass, an astronomer at the University of Arizona, was analyzing the relationship between sunspots and climate. He speculated that any change in solar energy caused by sunspot activity would cause changes in the amount of energy received by the Earth, and therefore would affect Earth's climate. Because climate affects the growth of plants, Douglass reasoned that sunspot cycles should be reflected in the growth of trees. His subsequent studies established a link between tree-ring widths and climate, and he soon confirmed the cross-matching of tree-ring widths among different trees, a technique known as crossdating. In 1914, Clark Wissler of the American Museum of Natural History noted Douglass's successful attempts at dating tree rings, and asked him if he could use tree-ring techniques to date the Puebloan ruins of the American Southwest. Fifteen years later, dendrochronology matured into an accepted science when Douglass overlapped his chronology developed from living trees with chronologies developed from trees used in many southwestern archeological sites. This accomplishment, published in *National Geographic Magazine* in 1929, provided archeologists with dates to the exact year when many southwestern pueblos were built. The importance of dendrochronology eventually led to the establishment of numerous tree-ring laboratories throughout the United States and Canada, Europe, South America, Asia, and Australia. *See* CLIMATOLOGY; DATING METHODS.

Annual ring formation. A tree grows by forming a new sheath or layer of woody tissue each year, much like a stack of cones one on top of the other. This growth occurs in a thin layer of cells that completely shrouds the tree just inside the bark, and is called the vascular cambium. Here, cambial cells divide, with outer cells contributing to the formation of phloem, and inner cells contributing to the formation of woody tissue. In temperate and subpolar regions, most trees break from winter dormancy and begin growth by forming new cells of wood using nutrient reserves stored from the previous growing season. In conifer trees, these cells are large and less dense, and have thin walls, producing light-colored wood called earlywood. Toward the end of the growing season, before the tree begins its dormancy period, smaller, denser, thick-walled cells are formed that are darker in color and called latewood. Taken together, these two bands of wood form an annual tree ring. In certain hardwoods, such as oaks, large cells called vessels are formed in the earlywood, while the latewood often lacks such vessels. *See* WOOD ANATOMY.

The formation of annual rings holds true for most trees in temperate and subpolar latitudes, but recent studies have shown some tropical and subtropical tree species also may produce annual rings. Occasionally, a tree may form latewood toward the end of the growing season, but a new flush of growth

due to favorable environmental conditions may again form earlywood cells. This intraannual band of latewood between earlywood bands is a false ring, but often is anatomically different from a true tree ring when magnified. In contrast, environmental conditions during some years may be so unfavorable that some trees may not form new wood, resulting in a missing ring, or a locally absent ring in the trunk of the tree. In most cases, the locations of false and locally absent rings are identified by dendrochronologists using the process of crossdating.

Principles. To search for and collect tree-ring data, dendrochronologists are guided by the following principles.

Limiting factors. Rates of plant processes are constrained by the primary environmental variable that is most limiting. For example, precipitation is often the most limiting factor to plant growth in arid and semiarid areas. In these regions, tree growth cannot proceed faster than that allowed by the amount of precipitation, causing the width of the rings (that is, the volume of wood produced) to be a function of precipitation.

Ecological amplitude. Plant processes function at their optimum within a defined geographic range constrained by one or more interacting limiting factors. Species are more sensitive to changes in environmental factors along the periphery of this geographic range. Tree species useful for dendrochronology are often found near the margins of their natural range.

Site selection. Sites useful to dendrochronology can be identified based on criteria that will produce tree-ring series sensitive to the environmental variable being examined. For example, trees especially responsive to drought conditions are usually found where the water supply is limiting, such as rocky outcrops and cliffs, mountain slopes, and arid and semiarid regions (**Fig. 1**). *See* PLANT-WATER RELATIONS.

Replication. The strength of the environmental signal being investigated can be maximized by collecting more than one sample per tree and more than one tree per site. Obtaining numerous trees from one site, and perhaps several sites in a region, ensures that the amount of noise (environmental or physiological factors not being studied) in a tree-ring series is minimized.

Collecting tree rings. After appropriate sites and trees have been located, dendrochronologists collect tree-ring data using an increment borer, a threaded hollow metal tube that is screwed into the trunk of a tree (**Fig. 2**). An extractor is inserted under the tree core that now resides inside the borer shaft, and is gently pulled from the tree. The hole left by the increment borer rarely does harm in conifer trees, but may cause internal wood discoloration and the formation of fungi that decay wood in hardwood tree species. Usually, at least two cores are extracted from opposite sides of the tree, and are placed in paper straws or plastic tubes to prevent breakage. In the laboratory, the pencil-thin cores are glued to grooved wooden mounts and sanded with progressively finer sandpaper (usually with a belt or orbital sander), until a high polish is obtained and the cell structure of the wood is easily discernible under a microscope. Occasionally, certain projects require complete cross sections from the trunks of dead trees, and a chainsaw is most often used.

Crossdating. Crossdating is the primary guiding principle in dendrochronology, and concerns the matching of patterns of ring widths from one tree with corresponding patterns for the same years from another tree (**Fig. 3**). Crossdating allows scientists to accurately assign calendar dates to tree rings by matching the sequence of tree-ring widths against a known reference chronology. Crossdating is possible because climate is largely a regional phenomenon, affecting trees in a like manner, so that similar patterns of ring widths are produced among many trees. A unique pattern of wide and narrow rings formed during a 50-year period is unlikely to be formed during any other 50-year period, because climate varies from year to year. Furthermore, crossdating helps identify false and locally absent rings that may otherwise be recorded as true rings.

In many laboratories, crossdating is accomplished with graphical comparisons called skeleton plots (**Fig. 4**) that accentuate the importance of narrow rings. The pattern of narrow rings is matched, or crossdated, against a reference chronology until a firm match is found. The tree-ring widths are then measured, and the crossdating accuracy is statistically confirmed using computer software. To ensure a high level of confidence in the dates assigned to tree rings, dendrochronologists look for high levels of statistical significance, higher than those normally used in statistics to identify a probable match. Only after a tree-ring series has been crossdated precisely,

Fig. 2. Small core being taken from a tree with an increment borer. (*Laboratory of Tree-Ring Research, University of Arizona*)

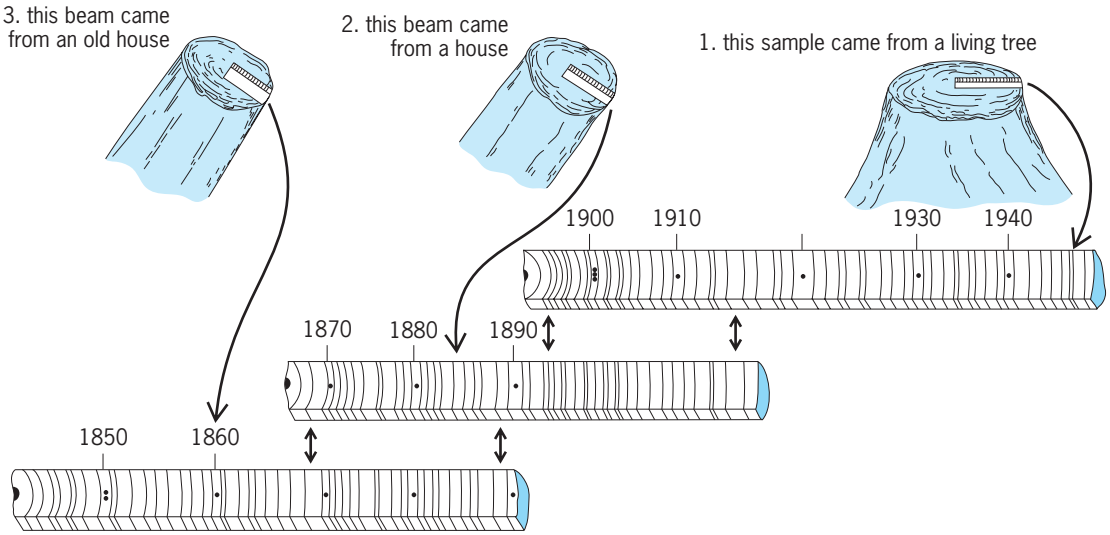

Fig. 3. Chronology building. Chronologies are extended backward in time by matching patterns of wood samples whose ages overlap.

both graphically and statistically, can dates for a wood sample be used.

Measuring tree rings. The widths of individual tree rings are measured using a measuring stage accurate to a hundredth of a millimeter. This stage is interfaced to a microscope, screen display, and a computer running software that records the measurement of each ring. Other physical properties of tree rings can be measured and used, including the average density of the tree ring, the minimum density of the earlywood, or the maximum density of the latewood, as well as the widths of the earlywood or latewood themselves. Density measurements are obtained using x-ray images made of tree cores that have been thin-sectioned with a microtome. In recent years, scanning techniques, image analysis, and sophisticated software have provided new techniques to easily capture tree-ring data. The information from scanned images of tree rings produces data similar to the changing density of the individual cells that constitute the tree rings.

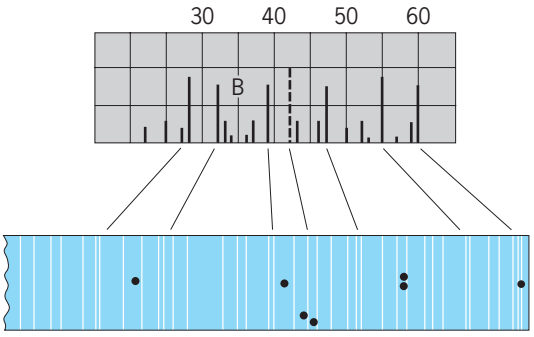

Fig. 4. Skeleton plot from an idealized tree-ring sequence. Ring widths are represented by vertical lines in an inverse proportion; that is, tall lines represent narrow rings. An exceptionally wide ring is designated by a B (big). The location of a missing ring, marked on the specimen by offset pinholes, is identified by a broken line. The number scale indicates years. (*After R. Berger, ed., Scientific Methods in Medieval Archeology, UCLA Center for Medieval and Renaissance Studies Contribution, IV, University of California Press, 1970*)

Chronology building. The measurements for individual years create a continuous time series for the full length of the individual cores. However, the raw measurements are rarely suitable for developing a site chronology because (1) trends are apparent simply due to physiological aging, (2) individual trees experience differences in growth rates, and (3) possible localized ecological disturbances (for example, the death of an individual tree) may impact growth of a subset of the sampled trees. Dendrochronologists are interested in average growth rates of trees at a site and must control for variation not due to climate. To accomplish this, all measurement series are separately standardized using complex mathematical expressions (for example, negative exponential curves) that provide predicted values of tree growth in any given year. The actual measurement is then divided by the predicted value (or sometimes the predicted value is subtracted from the actual measurement) to create an index of tree growth. The indices from all series for any given year are then averaged together to develop an average index. The final tree-ring index chronology represents the collective information from numerous trees within a site, and captures the majority of variability in tree-ring widths for a particular year.

Applications. Dendrochronology has become a useful tool in many areas of research.

Dendroarcheology. This science uses tree rings to date wood material from archeological sites or artifacts, and is most often applied in the southwestern United States and Europe. Dendroarcheology has established the exact year of construction for individual rooms in many ancient Native American ruins throughout the Southwest, providing important

Fig. 5. Bristlecone pines, which attain ages of more than 4000 years, form the basis for the longest chronology yet developed. (*Laboratory of Tree-Ring Research, University of Arizona*)

clues on changes in human populations hundreds of years before Euro-American settlement. In Europe, dendrochronologists often date the period of construction for barns, manors, cathedrals and churches, Roman bridges, wells and fountains, pile dwellings in lakes, and Neolithic settlements. Recently, dendrochronologists announced the development of a multimillennial tree-ring chronology that helped clarify Mediterranean archeological history. *See* ARCHEOLOGY.

Dendroclimatology. This science was first developed by Edmund Schulman, who was a student of Douglass and reconstructed climate for the American Southwest. Because temperature, precipitation, and other climatic variables affect tree growth, the climatic information can be mathematically extracted from the tree-ring record and reconstructed back in time for the length of the tree-ring record. For example, bristlecone pines growing in the White Mountains of eastern California have provided information on both temperature and precipitation fluctuations for the past 8000 years (**Fig. 5**). Reconstructions of annual rainfall for the American Southwest over the past 2000 years have provided new clues concerning century-scale climate change and its probable impact on ancient Native American cultures. Dendroclimatologists also use tree-ring records to quantify the rising levels of atmospheric carbon dioxide to better understand global warming.

Dendroecology. This science analyzes changes in ecological processes over time using tree-ring information. Forest trees are subject to a variety of ecological processes and disturbances that may alter forest structure and composition. For example, periodic insect outbreaks (such as the eastern spruce budworm) can defoliate or kill large stands of trees. Defoliation can be recognized in the tree-ring record as a group of continuous and progressively narrower tree rings. Other applications that fall under dendroecology include analyzing (1) the effects of air, water, and soil pollution on tree growth and forest health, (2) the age, maturity, and successional status of forest stands, and (3) the effects of human disturbances and management on forest vitality. *See* ECOLOGY.

Dendropyrochronology. This science reconstructs the history of wildfires from tree rings. Humans have greatly affected the natural course of wildfires, largely by fire suppression, often not realizing the beneficial nature of wildfires to forest ecosystems. In recent years, many land management agencies have sought to restore fire. This field takes advantage of the fact that a low-intensity surface fire will kill a part of the living cambium on the lower portion of the tree trunk which subsequent growth will preserve, thus leaving a fire scar in the tree-ring record. Dendrochronologists date tree rings that contain these fire scars to learn the frequency of past fires before widespread and pervasive human disturbances (such as fire suppression) occurred.

Dendrogeomorphology. This science studies earth surface processes using tree-ring data. Mass movements, such as landslides and snow avalanches, can kill or bend the trees in the movement path or along the edges of the movement. Tree-ring scientists can date when these trees were killed (by dating the outer ring of the tree) or when they were bent (by analyzing when dramatic changes in tree growth occurred). Therefore, a chronology of past landslide or avalanche events can be developed back in time to help assess the hazard potential for the low-lying areas at the base of these slopes. Tree rings are also used to learn about glacier movements and past climate by dating the moraines created along the sides and at the terminus of these glaciers. *See* GEOMORPHOLOGY.

Dendrohydrology. This science uses tree-ring data to investigate and reconstruct hydrologic properties, such as streamflow and riverflow, runoff, and past lake levels. This field is closely aligned with dendroclimatology, and requires statistical calibrations of tree-ring data, such as ring width and density, with hydrologic properties recorded during the historical period. For example, estimated flow for the Colorado River using information from gaging stations has been reconstructed from tree-ring data, and has shown that the flow of the river has fluctuated greatly prior to the twentieth century. Riverflow amounts based on twentieth-century observations are therefore not an accurate indicator of potential riverflow in the future. *See* HYDROLOGY.

Dendrochemistry. This important, emerging field of dendrochronology analyzes the chemical composition of tree rings, especially the mineral elements. Because trees record events in the environment, changes in the amounts of naturally occurring or

human-manufactured chemicals are recorded in the tree-ring record. Trees can provide a temporal record of toxic pollutants, such as cadmium, cesium, sulfur, and aluminum, to better understand the impact of pollution in the air, surface-water, or groundwater supply in ecosystems. The record of such trace elements is obtained by a variety of analytical methods that include neutron activation analysis, proton-induced x-ray emission, and atomic absorption spectrometry. *See* ANALYTICAL CHEMISTRY.

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Dendrology

The division of forestry concerned with taxonomy of trees and other woody plants. Dendrology, called forest botany in some countries, usually is limited to taxonomy of trees but may also include shrubs and woody vines. This basic subject in the training of foresters teaches how trees are named (nomenclature), described (morphology), and grouped (classification); how to find the name of an unknown tree and recognize important forest species (identification); and where trees occur both by geographic ranges of species and by forest types (distribution).

In forestry a tree is defined as a woody plant with a single erect perennial stem or trunk at least 3 in. (7.5 cm) in diameter at breast height (4½ ft or 1.4 m above the ground), a more or less definitely formed crown or foliage, and a height of at least 13 ft (4 m). A shrub is a woody plant, generally lower growing than a tree, and frequently having several slender perennial stems arising at or near the ground.

Common names of trees and lumber in the United States have been standardized by various forest agencies. The scientific name of a tree is in Latin and consists of two parts, a genus and a species; for example, *Pinus ponderosa*, ponderosa pine.

Trees are described and distinguished in botanical terminology largely in terms of their characteristics of form and structure (morphology). The principal parts useful in identification are leaves, flowers, fruits, seeds, buds, twigs, branches, trunk, bark, and wood. See separate articles on trees listed under common names.

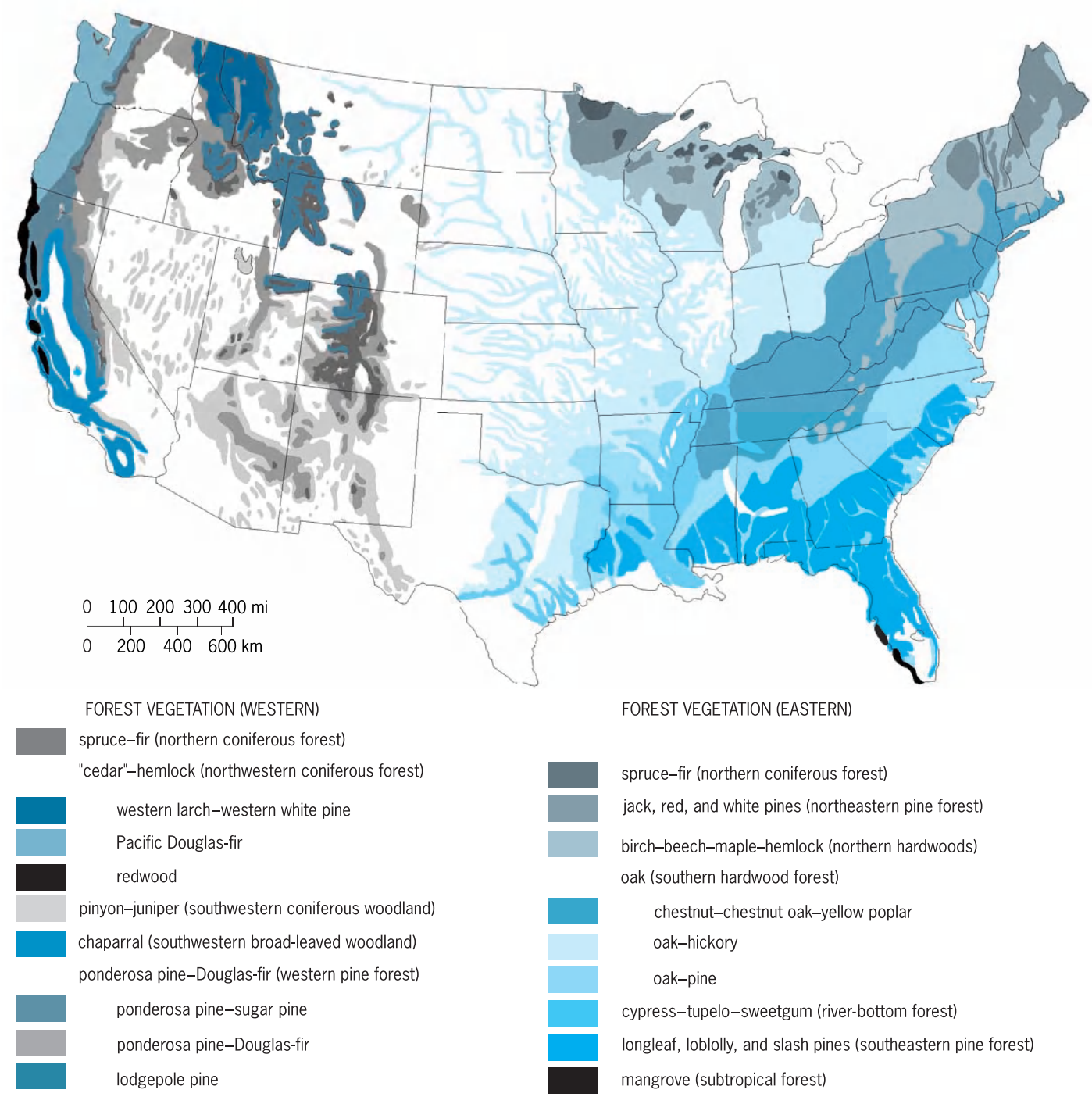
Classification. A common but artificial classification groups plants into trees, shrubs, and herbs. However, the 50,000 or more tree species in the world, as well as the numerous other species making up the plant kingdom, are arranged scientifically according to natural relationships as indicated by evolutionary evidence. For example, 10 genera, having a number of characteristics in common, compose the pine family (Pinacea). These and trees of related families make up the gymnosperms (Pinopsida), generally referred to as softwoods by lumber workers. Most tree species, however, are flowering plants, angiosperms (Magnoliophyta), and nearly all of these are dicotyledons (Magnoliopsida), often called hardwoods. Palms and bamboos are monocotyledons (Liliopsida). *See* PLANT ANATOMY; PLANT KINGDOM.

Native tree species in the continental United States total approximately 679 species in 216 genera and 73 families; 69 species are naturalized. However, only about 175 species in 50 genera are commercially important for lumber. About 100 are subtropical and tropical species of southern Florida. Many others are of limited occurrence, small size, or low-quality wood. Hawaii, with its unique tropical flora, has about 300 native tree species, nearly all not occurring elsewhere.

Identification. The correct scientific name of a recorded species may be determined by means of printed keys or manuals or by comparison with a known tree or with mounted specimens in a herbarium. Even in winter, leafless trees usually can be identified from keys by studying the bud, twig, and bark characteristics. Nearly every state publishes a popular, inexpensive, illustrated pocket guide or bulletin for the identification of the trees of that state. Regional floras are covered in other publications. *See* PLANT KEYS.

Distribution. Each tree species has its own natural distribution or range. A few tree species of the United States extend across the continent, while others are local and rare in distribution. About two-thirds of the important forest trees are eastern, or southeastern, whereas one-third are western, some in the Rocky Mountains, and others in the Pacific Coast region. Trees also have an altitudinal distribution and zonation in high mountains. *See* FOREST ECOSYSTEM.

Forest stands of similar composition, appearance, and structure are grouped together into areas characterized by major forest types or formation, and are named from the predominant or characteristic species (see **illus.**). The Society of American Foresters defined 90 forest cover types of eastern



Areas characterized by major forest types in the United States. (*U.S. Forest Service*)

North America (exclusive of Mexico) and 55 of western North America. *See* CHAPARRAL; FOREST AND FORESTRY; FOREST MANAGEMENT; FOREST MANAGEMENT; PLANT TAXONOMY. Elbert L. Little, Jr.

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Density

The mass of an object divided by its volume. Density is conventionally represented by the Greek letter rho: ρ. The most common units for density are g/cm^3 or specific gravity, which is dimensionless. Specific gravity for liquids or solids is usually defined as the density of a material divided by the density of water at 4°C (40°F), which is 1.0000 g/cm^3. Liquid water has a density close to 1 g/cm^3 at any reasonable

TABLE 1. Selected gas densities.

Material*	Temperature, °C (°F)	Density, kg/m³
Hydrogen	−252 (−422)	1.440
Hydrogen	20 (68)	0.090
Helium	20 (68)	0.179
Water (100°C)	100 (212)	0.588
Methane	20 (68)	0.716
Dry air	20 (68)	1.204
Oxygen	20 (68)	1.428
Sulfur dioxide	20 (68)	2.858
Chloroform	62 (144)	4.341
Xenon	20 (68)	5.857
Sulfur hexafluoride	20 (68)	6.072
Uranium hexafluoride	57 (135)	13.00

*In gaseous state, at a pressure of 1 atm.

temperature [it drops to 0.958 g/cm³ at 100°C (212°F), or boiling]. In U.S. customary units, at 40°F, a gallon of water weighs 8.345 lb (3.785 kg), while a cubic foot of water weighs 62.43 lb (28.32 kg). In modern SI units, 1 g/cm³ is frequently expressed as 1000 kg/m³. There are also a variety of archaic, industry-specific systems in which density is measured in degrees. An example of this is °API, still used in the petroleum industry. Degree measures of density are generally used as quantitative expressions of quality or composition, not as explicit measures of density. *See* HYDROMETER; MASS.

Gases. The density of a typical gas is about one thousandth that of a typical liquid, or about 0.001 g/cm³. However, the density of gases (**Table 1**) varies over a much wider range than that of liquids and solids. Gas density is dependent on three key factors: the material, the pressure, and the temperature. The dominant factor is pressure.

As pressure approaches zero (a vacuum), the density of any gas approaches zero. As pressure is increased, the density increases in one of two ways. At temperatures below the critical point, the gas will begin to condense into liquid or solid form at sufficiently high pressure. Liquids and solids are always considerably more dense than gases, so condensation is accompanied by a rapid increase in density. At temperatures above the critical point, the gas will simply increase in density indefinitely, without liquefying or solidifying. Each material has its own critical point: for water it is 374°C (705°F), while for nitrogen (N_2) it is 126 K (−147°C or −233°F).

At low pressures, the density ρ of most gases is very well approximated by the ideal gas law, Eq. (1),

$$\rho = \frac{MP}{RT} \tag{1}$$

where M is the molecular weight of the gas, P is the pressure, R is the ideal gas constant, and T is the absolute temperature. This equation is simplest in SI units, where ρ is in kilograms per cubic meter (kg/m³), M is in grams per mole (g/mol), P is in kilopascals (kPa), T is in kelvins (K), and R has the fixed value 8.3145 joules per mole per kelvin [J/(mol · K)]. For example, the molecular weight of air is about 29 g/mol, so that air at 293.15 K (20°C or 68°F) and atmospheric pressure (101.3 kPa) has a density of 1.2 kg/m³, or 0.0012 g/cm³. *See* METRIC SYSTEM; PHYSICAL MEASUREMENT; UNITS OF MEASUREMENT.

At higher pressures, it is necessary to account for interactions between molecules, and this is especially complicated near the critical point, where the densities of gases and liquids are comparable. There are several empirical correlations available for determining the density of nonideal gases; a particularly popular one is the van der Waals equation. The study of gas behavior under high pressure or near-critical conditions is an important part of the science of thermodynamics. *See* CRITICAL PHENOMENA; GAS; INTERMOLECULAR FORCES; VAN DER WAALS EQUATION; VIRIAL EQUATION.

Liquids. Liquid densities are more difficult to predict than those of gases. The density of a typical liquid is 1 g/cm³, although organic solvents often have lower density (octane: 0.703 g/cm³ at 20°C or 68°F) while liquid metals generally have much higher densities (mercury: 13.534 g/cm³ at 25°C or 77°F). The densities of liquids are far more dependent on the particular material than temperature or pressure (**Table 2**). Almost all liquids increase in density with decreasing temperature, but the effect is much more moderate than for gases. Over small ranges of temperature, the density of liquids is well represented by Eq. (2), where β is the thermal expansion coefficient, also known as the coefficient of cubic expansion, and ρ_0 is the density at the temperature T_0. An alternative expansion coefficient, the coefficient of linear expansion, α, is related by $\beta = 3\alpha$. The thermal expansion coefficient for most organic solvents is on the order of $\beta = 0.001\ \mathrm{K}^{-1}$. Over larger ranges of temperature, densities of liquids are more difficult to predict. Almost all liquids solidify at some point as temperature is decreased. The density generally increases significantly at the point of solidification. *See* THERMAL EXPANSION.

$$\rho = \rho_0[1 + \beta(T - T_0)] \tag{2}$$

Water is the most familiar liquid, but it is anomalous in terms of density. Water has a maximum density at 4.0°C (39.2°F). Below that temperature, water

TABLE 2. Selected liquid densities.

Material*	Temperature, °C (°F)	Density, g/cm³
Hydrogen	−253 (−423)	0.071
Nitrogen	−195 (−319)	0.807
Octane (gasoline)	20 (68)	0.703
Ethanol (ethyl alcohol)	20 (68)	0.789
Water	20 (68)	0.997
Glycerol (glycerin)	20 (68)	1.264
Chloroform	20 (68)	1.484
Perchloroethylene (PCE)	20 (68)	1.623
Lithium	181 (358)	0.534
Gallium	30 (86)	6.095
Lead	328 (622)	10.66
Mercury	20 (68)	13.55

*In liquid state, at a pressure of 1 atm.

TABLE 3. Selected solid densities.

Material*	Temperature, °C (°F)	Density, g/cm³
Paraffin wax	20 (68)	0.90
Ice (water)	0 (32)	0.92
Sodium chloride (salt)	20 (68)	2.17
Silicon dioxide (quartz)	20 (68)	2.65
Lithium	20 (68)	0.53
Aluminum	20 (68)	2.70
Titanium	20 (68)	4.54
Stainless steel	20 (68)	7.9
Copper	20 (68)	8.96
Lead	20 (68)	11.35
Gold	20 (68)	19.3
Osmium	20 (68)	22.57

*In solid state, at a pressure of 1 atm.

actually decreases in density with decreasing temperature. Furthermore, when water solidifies into ice (at 0°C or 32°F) it decreases in density significantly. At atmospheric pressure and 0°C, the density of liquid water at freezing is 0.9998 g/cm^3 while the density of solid ice is 0.9167 g/cm^3. Most liquids freeze from the bottom; the buoyancy of ice makes frozen lakes and oceans possible without complete freezing. The density of water is often modified by soluble or particulate content. Typical ocean water (15°C or 59°F) has a density of 1.026 g/cm^3, while normal human blood (37°C or 98.6°F) has a density of about 1.060 g/cm^3 (it contains salts, biological materials, and red blood cells). At low concentrations, the change in density (from pure water) is directly proportional to the concentration of soluble or particulate content. *See* LIQUID; SEAWATER; WATER.

Solids. Solid densities (**Table 3**) are similar to those of liquids, except that solids are generally slightly denser, and solid densities are somewhat less dependent on pressure and temperature. The density of nonporous solids is primarily dependent on the particular material, and secondarily on the structure, which may be amorphous or crystalline. Transitions in solid structure are generally accompanied by significant changes in density. For example, diamond and graphite are both forms of elemental carbon; the density of diamond is 3.2–3.5 g/cm^3, that of graphite, 1.9–2.3 g/cm^3. For solids of fixed structure, Eq. (2) may be used over fair ranges of temperature. Typical thermal expansion coefficients are $\beta = 5 \times 10^{-5}\ \mathrm{K}^{-1}$ for copper (over a broad range of temperatures), and $\beta = 2 \times 10^{-4}\ \mathrm{K}^{-1}$ for ice (at 0°C). When thermal expansion coefficients are mismatched between adjacent solids, extreme forces (thermal stresses) can be generated by changes in temperature. This is the reason why small gaps are typically left between different construction materials, and why vinyl siding is usually only loosely fixed to wooden houses. *See* THERMAL STRESS.

Many solids are commonly encountered in granular or porous form (soils, for example). The density of such materials is difficult to define when the granules or pores are too small to be easily distinguished from the bulk material. Many modern materials incorporate features at the molecular scale (zeolites and nanotubes, for example). The density of the nonporous solid is generally significantly different from porous materials infused with air (such as powders and foams) or water (such as slurries and clays). *See* CARBON; NANOSTRUCTURE; ZEOLITE.

The most dense ordinary solids (and liquids) are heavy metal elements and their alloys. The densest elements are the metals osmium and iridium, with densities of 22.6 and 22.4 g/cm^3 (at 20°C), respectively. The least dense metal is lithium, at 0.53 g/cm^3 (at 20°C). The least dense solids are a kind of solgels known as aerogels, generated by evaporating off a solvent from a gelatinous liquid solution. One common type of aerogel, based on silica, can have a density as low as 0.003 g/cm^3. The low density of aerogels makes them excellent insulators. *See* GEL.

Extremes. While the most extreme densities found on Earth are air and heavy ores, interstellar space and stars exhibit a far broader range of density. The density of open space is practically zero, but there are measurable amounts of matter even in intergalactic space (the exact nature of such matter is currently the topic of much debate). At extremely high densities, quantum-mechanical effects become very significant. Densities in ordinary stars such as the Sun are not extreme by terrestrial standards, despite the extreme pressures and temperatures. The mean density of the Sun is only 1.4 g/cm^3. However, stars eventually collapse into white dwarfs, neutron stars, or even black holes. The densities of a typical white dwarf and neutron star are 10^6 and 10^{15} g/cm^3, respectively. *See* BLACK HOLE; NEUTRON STAR; SUN; WHITE DWARF.

Measurement. Density measurement devices are commonly referred to as densitometers, and density measurement is sometimes referred to as densitometry. The American Society for Testing and Materials (ASTM) defines a broad range of standard tests for measuring density, which should be used wherever standardization is critical. There are a wide variety of handbooks with extensive data on the densities of common substances and materials.

Most modern density measurement techniques are indirect, as weighing materials tends to be imprecise. However, certain liquids and solids are easily weighed and have well-defined volumes. Liquid density can generally be determined by placing a sample of fluid in a volumetric flask of known volume and weight; the difference between the weight of the flask with and without the liquid divided by the volume gives the density directly. Nonporous solid volumes are easily determined using Archimedes' method: Place the solid in a full container of water or another liquid, and measure the volume displaced (overflow), which is the volume of the solid. The mean solid density follows from the weight. *See* ARCHIMEDES' PRINCIPLE.

Where gas densities cannot be predicted, they are typically measured via momentum. Bernoulli's law states that flowing gas—through a tube, venturi (**Fig. 1**), orifice, or turbine (**Fig. 2**)—generates a pressure difference that is directly proportional to

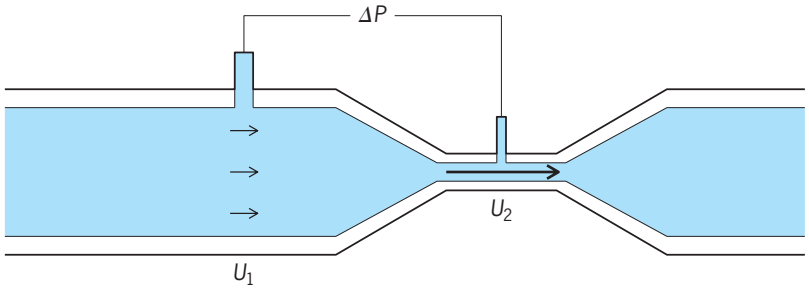

Fig. 1. Venturi flow constriction, used to measure gas or liquid density, showing ports for measurement of the pressure difference ΔP, which is determined by the density and the flow velocities, U_1 and U_2, at the ports.

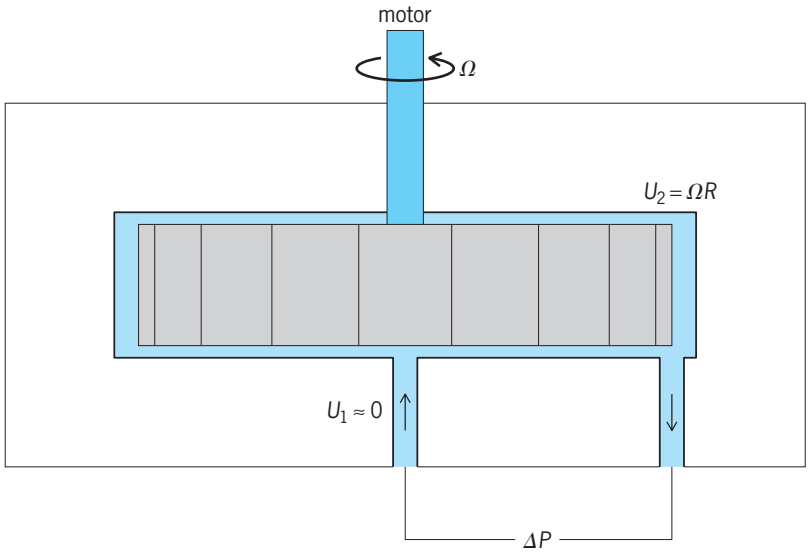

Fig. 2. Turbine driven by motor, used to measure gas density, showing ports for measurement of the pressure difference ΔP. This quantity is determined by the density and the flow velocities, $U_1 \approx 0$ and $U_2 = \Omega R$, at the ports, where R and Ω are the radius and angular velocity of the turbine.

the density of the fluid. It is given by Eq. (3), where

$$\Delta P = \rho \left(U_1^2 - U_2^2 \right) = K\rho \tag{3}$$

ΔP is the measured pressure difference, U_1 and U_2 are flow velocities (normally held constant for density measurements), and K is a calibration constant. Devices with turbines are especially convenient as they can act as a measurement device and a pump simultaneously. Gas densities are sometimes expressed as specific gravity with respect to dry air at a specific temperature, as air is a convenient calibration material with a well-defined density. Liquid density can also be measured via pressure differences, but many viscous liquids only follow Bernoulli's law [Eq. (3)] only at prohibitively high flow rates. *See* BERNOULLI'S PRINCIPLE; FLOW MEASUREMENT; TURBINE.

Liquid density can be measured very precisely using buoyancy effects. Objects immersed in liquid behave as if their density has been reduced by that of the liquid: $\rho_{\text{effective}} = \rho - \rho_{\text{liquid}}$. Astronauts use this principle in special pools where they are neutrally buoyant (same density as the pool liquid), which mimics zero gravity well. Hydrometers (**Fig. 3**) are flotation devices in which the level of the float above the liquid surface gives density directly. Other buoyancy-based devices measure forces on a float, where the force is zero when the float and liquid densities are equal. *See* BUOYANCY.

The density of solids and liquids is often measured via absorption of radiation. For materials of identical atomic composition, the absorption rate of x-rays or gamma rays is proportional to the density of the material. One advantage of radiation densitometers is that they can be mounted external to pipes or containers. Disadvantages include the clearances and safety precautions normally required to use radioactive materials in practice. *See* RADIOACTIVITY AND RADIATION APPLICATIONS.

Vibrating densitometers make use of the fact that springs and other mechanisms vibrate with periods proportional to the square-root of mass. For a fixed volume of material in a container of fixed mass, the period of vibration is given by Eq. (4), where P is the

$$\rho = AP^2 - B \tag{4}$$

period, and A and B are calibration constants. There are many types of vibrating densitometers, and most have a relatively broad range of measurement with a high degree of precision. Unfortunately the calibration constants tend to be dependent on temperature and can also drift with time. *See* VIBRATION.

Many materials are simple in form but complicated in composition. Many common materials bond

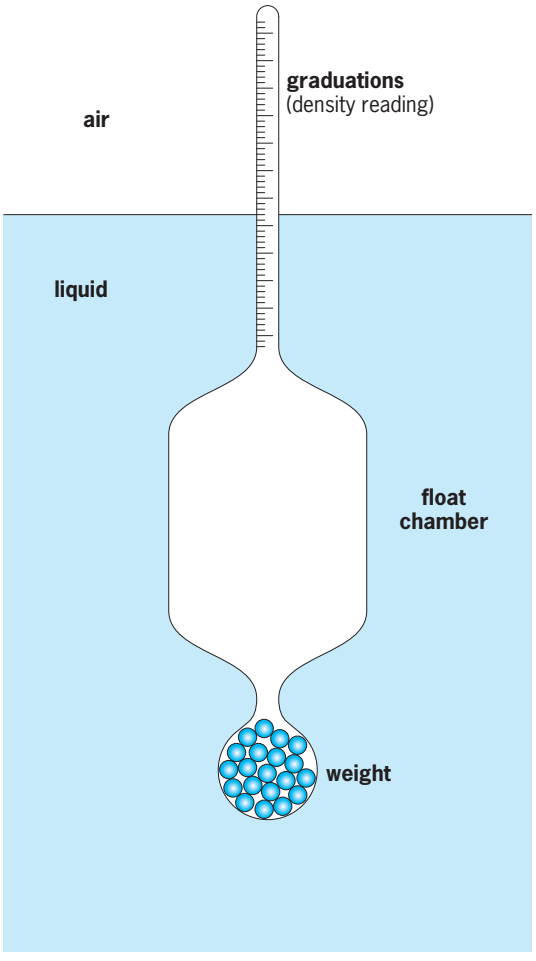

Fig. 3. Hydrometer.

chemically with water, as well as absorbing additional amounts into their structure. Wood, for example, can contain significant amounts of water without appearing wet; some of this is bound chemically while the rest is held physically. The density of wood and other wet materials is generally highly dependent on water content, and "dry" materials are typically defined as materials where all of the physically bound water has been removed (but not the chemically bound water). Porous or contaminated materials pose additional challenges. Often pores must be cleared by washing with water or other solvents and then evaporating off the cleaning material. Cleaning generally changes the density of a material in numerous ways, by dissolving soluble salts or collapsing pores, for example. The cleaning material may also associate chemically or physically with the solid, and must be chosen carefully to avoid contamination. Sol-gels are especially prone to modification; silica-based aerogels collapse when exposed to liquid water or even high humidity. *See* HYDRATE; HYDRATION.

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Density matrix

A matrix which is constructed as the most general statistical description of the states of a many-particle quantum-mechanical system. The state of a quantum system is described by a normalized wave function $\psi(x,t)$ [where x stands for all coordinates of the system, and t for the time], which satisfies the Schrödinger equation (1), where H is the hamiltonian of the system, and $\hbar$ is Planck's constant divided by 2π. Furthermore, $\psi(x,t)$ may be expanded in terms of a complete orthonormal set $\{\varphi(x)\}$, as in Eq. (2). Then, the density matrix is defined by Eq. (3), and this density matrix describes a pure state.

$$H\psi(x,t) = \frac{\hbar}{i}\frac{\partial\psi(x,t)}{\partial t} \tag{1}$$

$$\psi(x,t) = \sum_n a_n(t)\varphi_n(x) \tag{2}$$

$$\rho_{mn}(t) = a_n^*(t)a_m(t) \tag{3}$$

Examples of pure states are a beam of polarized electrons and the photons in a coherent beam emitted from a laser. *See* LASER; NONRELATIVISTIC QUANTUM THEORY.

In quantum statistics, one deals with an ensemble of N systems which have the same hamiltonian. If the αth member of the ensemble is in the state ψ^α in Eq. (4), the density matrix is defined as the ensemble average, Eq. (5). In general, this density matrix describes a mixed state, for example, a beam of unpolarized electrons or the photons emitted from an incoherent source such as an incandescent lamp. The pure state is a special case of the mixed state when all members of the ensemble are in the same state. *See* STATISTICAL MECHANICS.

$$\psi^\alpha(x,t) = \sum_n a_n^\alpha(t)\varphi_n(x) \tag{4}$$

$$\rho_{mn}(t) = \frac{1}{N}\sum_\alpha [a_n^\alpha(t)]^* a_m^\alpha(t) \tag{5}$$

Diagonal elements. The diagonal element $\rho_{mn}(t) = |a_n(t)|^2$ of the density matrix for a pure state gives the probability of finding the quantum system in the state $\varphi_n(x)$ at time t. The set $\{\varphi(x)\}$ is complete; property of the probabilities implies that $\Sigma_n\rho_{nn}(t) = \mathrm{Tr}$ unity also for a mixed state. Since all diagonal elements are nonnegative, it also follows that $0 \le \rho_{nn} \le 1$.

Observables. The ensemble average of an observable A (an operator that represents an observable quantity) can be written as Eq. (6). Since $\overline{\overline{A}}$ is the trace of a matrix, its value is independent of the choice of the orthonormal set $\{\varphi(x)\}$.

$$\overline{\overline{A}} = \frac{1}{N}\sum_\alpha\sum_{mn}\left(a_n^\alpha\right)^*\left(a_m^\alpha\right)\int \varphi_n^* A\varphi_m dx$$
$$= \sum_{mn}\rho_{mn}A_{nm} = \mathrm{Tr}\left(\rho A\right) \tag{6}$$

Pure and mixed states. For a pure state, Eq. (7) holds, or, equivalently, $\rho^2 = \rho$ or $\mathrm{Tr}\,\rho^2 = 1$. But, for a mixed state, $\mathrm{Tr}\,\rho^2 < 1$ always holds. The necessary and sufficient condition for a state to be a pure state is $\rho^2 = \rho$.

$$(\rho^2)_{mn} = \sum_p a_p^* a_m a_n^* a_p = a_n^* a_m = \rho_{mn} \tag{7}$$

Eigenvalues and eigenvectors. The density matrix is hermitian, $\rho^*_{mn} = \rho_{nm}$, and is diagonalizable. The eigenvalues of ρ for a pure state are all zero except one, which has the value unity. The eigenvector corresponding to the nonzero eigenvalue is the pure state the system is in. For a mixed state, the eigenvectors are the possible pure states the system may have, and the eigenvalues are the probabilities of these states. Thus all eigenvalues lie between 0 and 1. The extreme case of a mixed state is the random state for which all N members of the ensemble are in different states. All eigenvalues corresponding to the occupied states have the value $1/N$. *See* MATRIX THEORY.

Liouville equation. The density matrix satisfies the Liouville equation (8), where H is the hamiltonian of the system. In a state of equilibrium $\partial\rho/\partial t = 0$. Then, Eq. (8) implies that ρ is a function of the hamiltonian H alone. For an ensemble in thermal equilibrium, ρ is related to temperature

$$i\hbar\frac{\partial\rho}{\partial t} = [H,\rho] \tag{8}$$

T according to Eq. (9), where k is the Boltzmann

$$\rho = \frac{e^{-H/kT}}{\text{Tr } e^{-H/kT}} \tag{9}$$

constant. This result, together with Eq. (6), allows a complete determination of the equilibrium thermodynamic properties of the system. For example, the entropy S is proportional to the expectation value of $\ln \rho$ as given in Eq. (10).

$$S = -k\,\overline{\overline{\ln \rho}} = -k\,\text{Tr}\,(\rho\ \ln\ \rho) \tag{10}$$

The relationship in Eq. (10) between the entropy and the density matrix holds for nonequilibrium systems as well. On the basis of knowledge of the eigenvalues of the density matrix, it can be deduced that the entropy of a pure state is zero, that of a mixed state is nonnegative, and that of a random state has the highest possible value of $k \ln N$ for an ensemble of N systems. Hence, entropy is a measure of the randomness of the state of an ensemble. *See* ENTROPY.

Effect of measurement. Another consequence of the Liouville equation, Eq. (8), is that, for an isolated system whose hamiltonian is time-independent, a pure state cannot evolve into a mixed state, or vice versa, as time goes on. The act of a physical measurement, however, can interrupt the time development of a system so that a pure state can be turned into a mixed state. Consider a Stern-Gerlach experiment which employs a nonuniform magnetic field to separate a beam of electrons into two beams which have spins parallel and antiparallel to the field. The direction of the field is chosen as the z axis. The incoming electrons are assumed to be polarized along an axis different from the z axis. The wave function of the incoming beam is then $\psi = a\psi_+ + b\psi_-$, where ψ_+, ψ_- are the wave functions of the spin-up and spin-down states and $|a|^2 + |b|^2 = 1$. The density matrix is shown in Eq. (11). This matrix satisfies the requirement

$$\rho = \begin{pmatrix} |a|^2 & ab^* \\ a^*b & |b|^2 \end{pmatrix} \tag{11}$$

$\rho^2 = \rho$ for a pure state. After passing through the Stern-Gerlach apparatus, the density matrix becomes Eq. (12), which represents a mixed state. The measurement has resulted in an increase in the entropy or randomness of the electron beam.

$$\rho' = \begin{pmatrix} |a|^2 & 0 \\ 0 & |b|^2 \end{pmatrix} \tag{12}$$

Perturbations. The Liouville equation is also very useful in determining the effect of a perturbation on the density matrix and, hence, on the properties of the system. Let the unperturbed hamiltonian be H_0 and the perturbing hamiltonian be H_1. The density matrix is similarly written as the sum of the unperturbed part ρ_0 and a correction ρ_1. Furthermore, it is assumed that the system is in equilibrium state before the perturbation is applied. Then Eq. (8) takes the form of Eq. (13). If the perturbation is weak, the last term of Eq.(13) can be ignored because it is of the second order of the perturbation. When this is done, the correction term can be solved to obtain Eq. (14). The Heisenberg operator $H_1(t - t')$ is defined in Eq. (15). The lower limit of the integration in Eq. (14) is the time at which the perturbation is turned on. *See* PERTURBATION (QUANTUM MECHANICS).

$$i\hbar\,\frac{\partial \rho_1}{\partial t} = [H_0, \rho_1] + [H_1, \rho_0] + [H_1, \rho_1] \tag{13}$$

$$\rho_1 = -\frac{i}{\hbar}\int_{t_0}^{t_1} [H_1(t - t'),\ \rho_0]\,dt' \tag{14}$$

$$H_1(t - t') = \exp\left[\frac{i}{\hbar}H_0(t - t')\right]H_1 \cdot \exp\left[-\frac{1}{\hbar}H_1(t - t')\right] \tag{15}$$

From this result the first-order correction to the expectation value of an observable A can be calculated as in Eq. (16). Notice that ρ_1 involves only the perturbation and the equilibrium density matrix. Thus, the first-order correction, called the linear response of the system to the perturbation, can be completely calculated from the knowledge of the equilibrium state. This is the starting point of the modern theory of transport phenomena. *See* KINETIC THEORY OF MATTER.

$$\overline{\overline{A}}_1 = \text{Tr}\,(\rho_1 A) \tag{16}$$

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Dental anthropology

The scientific study of people, with their living and extinct primate relatives, using the evidence of teeth. Dental anthropologists include not only those trained in anthropology but also practicing dentists, anatomists, radiologists, forensic scientists, biochemists and geneticists, archeologists, paleontologists, and zoologists.

Dentition. The deciduous (milk) teeth and permanent (adult) teeth make up the dentition. They have their own distinctive patterns of development, physiology, size and shape, disease, and changes with age, all very different from the bones of the skeleton. In particular, the three tissues of the teeth—enamel, dentine, and cement—do not turn over continuously as do bone and most other tissues of the body. Once formed in childhood, the main structure of a tooth changes very little throughout adult life, although it is modified by wear or disease, and there is slow continued deposition of dentine within the pulp chamber, and of cement on the root surface. This means

that, even in adults, tooth form is more clearly related to the gene-mediated control of development than a structure such as a bone, which is modified throughout life in response to changing physiological factors and mechanical loadings. Many features of the tooth crown have a high heritability. Before modern molecular genetics, dental morphology was proposed as a simple method for establishing the relationship of twins. It has become possible to demonstrate the activity of particular genes in the differentiation of tooth form. Another effect of this conservation of original tooth structure is that the layered pattern of growth in dental tissues is preserved throughout life, so that childhood development can be studied even in adults. For archeologists and paleoanthropologists, there is a further advantage in the study of teeth—they are exposed in the mouth. Their adaptation to this harsh, abrasive environment in life ensures that they are the toughest part of the body, and survive relatively unchanged in a wide variety of burial environments. The majority of fossils are teeth and jaws. In addition, teeth were in contact with all the food eaten, and were used not only for chewing but for initial food preparation, grooming, acting as a third hand, and so on. Therefore, teeth bear many traces of the diet and life-style of their owners. Furthermore, teeth are easily studied in living people and animals by direct observation, by taking dental impressions (casts), or by examining teeth extracted in dental surgeries. Direct comparisons can be made far more easily than with bony fossils.

Morphology. For many dental anthropologists, the core of the subject is tooth morphology. Identification of different primate species is most readily done from tooth form in archeological and fossil material. One feature of human evolution has been a reduction in tooth crown size. A rapid dental reduction has taken place over the last 50,000 years in many parts of the world. This may be part of a general change in body proportions or may relate to particular dietary or technological changes. Among the higher primates as a whole, males tend to have larger teeth than females, particularly the canine teeth. Even modern humans show this to a small extent, and it helps to distinguish between the sexes in archeological and forensic remains. It does, however, complicate the definition of species from the fossils of extinct primates. The cusps, ridges, and furrows that decorate the crown surface also vary within different species of primates, together with the number and form of tooth roots. The different forms have characteristic frequencies for different populations of living people, and are used to investigate their relationships and history in comparison with archeological groups. Teeth additionally vary in their placement and in the way in which they fit together—the dental occlusion. Occlusal anomalies have become more common with the adoption of modern westernized life-styles and diet.

Development. The timing of formation has been established for the different types of teeth in the dentition from x-ray studies of living children, and of the young of several species of other primates. The dentition develops in a well-defined series of stages, so the state of dental development at death is an important guide to a child's age in archeological and forensic cases. It is more difficult to apply the concept to extinct primates, so attempts have been made to calibrate their development sequence by counting regular microscopic structures developing during the formation of enamel. This technique makes it possible to produce a development chronology for the dentition of one individual, providing such detail that episodes in which growth is disturbed (caused by factors such as childhood fevers or dietary deficiencies) are recorded as defects in the sequence. Estimation of age in the remains of adults is much more difficult, but teeth still provide important methods for forensic and archeological studies. Gradual microscopic changes take place in the dentine and cement, and the tooth crowns are progressively worn. Wear rate was much greater among people in the past than in most modern societies. Not only did the food involve more energetic chewing and contain more abrasive particles, but the teeth were used for a variety of other tasks. These factors leave their mark in a variety of anomalous wear patterns. At a microscopic level, the pattern of wear scratches on the enamel (microwear) has been used to compare the diet of different living primate species and to reconstruct the diet of extinct species. The nature of the diet also leaves its mark in the epidemiology of dental diseases. There are differences, for example, among the great apes in the frequency of dental caries (decay) in relation to the proportion of sugar-containing fruit in their diet. Among humans, hunter-gatherer groups typically had very low caries rates and, for example, there was a clear increase associated with the adoption of maize agriculture by native Americans. The much higher caries rates of the last two centuries are associated with high sugar consumption. Associated dental treatment leaves its mark, and one of the most important methods for personal identification in forensic cases is through the matching of this evidence with dental records. *See* DENTAL CARIES; DENTISTRY; TOOTH.

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Dental caries

A disease in which the mineralized tissues of the tooth undergo progressive destruction from the outside surface of the tooth. It is caused by bacteria that colonize the tooth surface and, under certain conditions, produce sufficient acids to demineralize the enamel covering of the tooth crown or the cementum covering the root, and then the underlying dentin. As the destruction of the dentin progresses,

along with breakdown of the organic components, the bacteria invade the dead tissue and enter the pulp chamber. The pulpal tissue becomes infected and the typical toothache may ensue. The infection can ultimately destroy the pulpal tissue and extend out through the apical openings of the roots and into the surrounding periodontal tissues.

Anatomy. The carious lesion begins on the surface of the enamel of the tooth crown. If the gum tissue has receded, exposing the surface of the root, the lesion may begin on the cementum surface of the exposed root. Since the bacteria that cause caries reside in dental plaque that forms on the tooth surface, sites that retain plaque are at the greatest risk for the development of carious lesions. Thus, dental caries most often develops in the pits and grooves or fissures of the occlusal (biting) surfaces of the premolars and molars, on the smooth surfaces of the crown just below the contact areas of the adjacent surfaces, and on the smooth surfaces of the facial surfaces (facing lips or cheeks) or lingual surfaces (facing the tongue) adjacent to the margin of the gum tissue. The progression of the carious lesion through dental enamel follows the direction of the enamel rods that extend from the surface of the crown to the dentinoenamel junction, that is, the junction of dentin and enamel. As a result, carious lesions originating in the occlusal pits and fissures of posterior tooth crowns have very small surface defects that expand in size as they progress toward the dentinoenamel junction. Those originating in the enamel of the smooth surfaces of the tooth are broadest at the surface and decrease in breadth as they approach the dentinoenamel junction. All carious lesions expand laterally at the dentinoenamel junction and progress within the dentin as a conical lesion with its apex directed inward.

One of the most important anatomical features of caries is the very early development of an enamel lesion, also referred to as the incipient lesion. As acid is produced in dental plaque next to the surface of the enamel, it penetrates the enamel between the crystals of mineral and begins to dissolve these crystals, liberating mainly calcium and phosphate ions. As these ions diffuse out of the enamel, the mineral reprecipitates in the surface layer because of the high calcium and phosphate content of the dental plaque overlying the area. This complex demineralization-remineralization process leads to the development of a lesion that is demineralized, but covered superficially by a relatively intact surface zone. Since the crystals of mineral within the partially demineralized subsurface zone are not totally dissolved, this lesion can remineralize in the absence of further acid attack. Hence, the early enamel lesion is regarded as a reversible condition. Trace amounts of the fluoride ion (tenths of a part per million) are known to catalyze the remineralization of incipient lesions, so it is now believed that this is one of a number of ways in which fluoride prevents dental caries. If the lesion progresses beyond the incipient stage and there has been cavitation with total loss of mineral crystals, the lesion is no longer reversible and the tooth must be restored with filling material such as plastic, amalgam, or gold.

Etiology and pathogenesis. The sites of caries development have been correlated with the presence of dental plaque containing one of a number of cariogenic streptococci referred to as mutans streptococci—notably in humans, *Streptococcus mutans* and *S. sobrinus*. These microorganisms are cariogenic because of their ability to generate a considerable amount of acid as a result of their metabolism of carbohydrates, and to survive in an acid environment. The acid environment is only slowly neutralized since its dilution and neutralization is hindered by the fact that the acid is within the dental plaque and thereby shielded from the saliva. Each time a carbohydrate-rich substance is ingested, acid is formed. Therefore, the frequency of ingestion and physical consistency of fermentable carbohydrates are important factors in caries formation. *See* STREPTOCOCCUS.

The progression of development of a carious lesion depends on the amount of time that a site on a tooth is subjected to acidic conditions (low pH) relative to the amount of time that the site is at neutral pH. If demineralization exceeds remineralization, the carious site progresses from the incipient lesion stage to the stage of cavitation where the surface layer of enamel is lost because of excessive mineral destruction. From this point, the lesion can increase in size and involve the underlying dentin, and finally the pulpal tissue. As the carious lesion approaches the pulpal tissue, bacterial toxins and other products gain access to this tissue and elicit an inflammatory reaction which intensifies as the bacteria themselves enter and infect the pulp. Inflammation of the pulp (pulpitis) is usually followed by death of this tissue (pulpal necrosis) and in most cases the infection extends through the apical foramina of the roots and into the adjacent periodontal ligament space.

Prevention. Dental caries can be prevented by making the tooth less susceptible to acid attack, removing cariogenic bacteria from the teeth, and limiting ingestion of cariogenic substrate.

The most effective and least expensive means of rendering the tooth less susceptible to the caries attack is through fluoridation, either by systemic means such as water fluoridation, or by topical application of fluoride to the tooth surfaces. Water fluoridation (approximately 1 ppm) by itself reduces the incidence of caries by 50–60%. Topical fluoridation can be instituted by many means, ranging from professional applications in the dental office to self-applied fluoride in the form of toothpastes, mouth rinses, and gels. The mode of action of fluoridation is not fully understood, but the benefit to the tooth structure is probably a result of a number of actions. It is believed that fluoride, when incorporated into the mineral phase of enamel and the other hard tissues, results in a better crystalline structure that is less susceptible to acid dissolution. Alternatively, the anticaries effect of fluoride may be due to fluoride's ability to inhibit demineralization and promote remineralization. In addition to fluoridation,

means of improving the tooth's resistance to dental caries include the application of plastic pit and fissure sealants which eliminate retentive areas on the biting surfaces of the posterior teeth that are more susceptible to carious lesions.

Cariogenic bacteria in the dental plaque can be reduced or eliminated by physical means such as daily toothbrushing and flossing. Although antibiotics are known to be effective against dental caries, the use of agents such as penicillin cannot be justified because of the danger of causing hypersensitivity to the antibiotic, or promoting the development of antibiotic-resistant strains that could be much more dangerous than the normal ones found in the mouth. The antiseptic chlorhexidine has been shown to be effective in reducing dental plaque and caries in children when used in the form of a mouth rinse.

Since cariogenic bacteria cannot form acid in the absence of fermentable carbohydrates, the risk of developing caries can be diminished significantly by controlling the dietary intake of carbohydrate, mainly sucrose. Not only is sucrose an excellent substrate with which plaque microorganisms form acid by-products, but it also promotes the accumulation of mutans streptococci on the teeth. Acid formation continues as long as there is fermentable carbohydrate available to the bacteria. *See* MICROBIOLOGY; TOOTH.

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Dentistry

The branch of biomedical science concerned with the embryology, development, and structure of the teeth, jaws, oral cavity, and adjacent structures, and with the prevention, diagnosis, and treatment of their abnormalities.

The practice of general dentistry requires a thorough knowledge of not only the structure, growth, function, and pathologic conditions of the oral cavity and related structures, but also the relationship of the oral cavity to other parts of the body in health and disease, and a recognition of oral manifestations of various systemic diseases. It is highly demanding because of the wide spectrum of diseases and abnormalities encountered. Nevertheless, most of the dentist's time and effort are devoted to the treatment of dental caries (decay) and the periodontal diseases. In the United States the prevalence of dental caries, especially in children and young adults, has decreased rapidly; a similar trend is occurring with gingivitis (inflammation of the mucous membrane surrounding the tooth sockets) and periodontitis (inflammation of the tissues surrounding the teeth). As a consequence, other conditions are becoming relatively more important in general dentistry. These include orofacial and temporomandibular joint pain; oral mucosal lesions associated with immunosuppresion, irradiation, and chemotherapy in patients with cancer or organ transplants; and unusual mucosal and periodontal lesions in patients with acquired immune deficiency syndrome (AIDS).

Restorative dentistry attempts to preserve and restore decayed, defective, missing, and traumatized teeth. In addition to biological and medical science, restorative dentistry calls upon engineering, metallurgy, and polymer chemistry, coupled with a high degree of manual dexterity. Restorative treatment can result in improved chewing, better appearance, and correction of some speech abnormalities; it can also alleviate joint and muscle pain. Advances include the development of composite resins strong enough to withstand biting and chewing; some of these can be bonded directly to the teeth.

The treatment of gingivitis and periodontitis requires the removal of bacterial deposits (calculus and plaque) from the crowns and roots of affected teeth (known as scaling and root planing) and adoption of hygienic practices by the patient. In some cases, especially in moderate to severe periodontitis, surgery is needed to reach diseased root surfaces or to eliminate periodontal pockets. Special care is required by patients with joint or cardiac valve replacements or vascular grafts or with a history of such systemic diseases as rheumatic fever, diabetes mellitus, hepatitis, or AIDS. *See* ACQUIRED IMMUNE DEFICIENCY SYNDROME (AIDS); CANCER (MEDICINE); DIABETES; HEPATITIS; RHEUMATIC FEVER.

Specialization

The complexity of modern methods of diagnosis and treatment has led to the recognition of eight branches of specialization by the American Dental Association: oral and maxillofacial surgery, endodontics, orthodontics, pedodontics, periodontics, prosthodontics, oral pathology, and public health dentistry. Subspecialties include oral medicine, dental radiology, oral diagnosis, and gerodontology.

Oral and maxillofacial surgery. This branch of dentistry treats diseases and abnormalities of the maxillofacial region by surgical means. Oral surgeons treat a wide variety of problems by removing teeth, cysts, tumors, and growths; reducing bone fractures; and correcting congenital anomalies and malformations of the maxillofacial region. Arthroscopy using an endoscope to examine the interior of a joint to diagnose for diseases of the temporomandibular joint, and replacement of missing teeth with implants are areas of importance.

Endodontics. Teeth contain a pulp chamber occupied by nerves, blood vessels, and connective tissue. The tooth pulp can die as a result of infected carious lesions or of trauma. Endodontists treat teeth that are nonvital or made so during restoration. With files and reamers, the pulp tissue is removed and root canals are enlarged, sterilized, and filled, usually with guttapercha (a heat-sensitive plastic substance). Following root canal therapy, the tooth dentin is without sensitivity, although the cementum (calcified tissue)

covering the tooth surface and attaching the tooth to the alveolar bone remains healthy.

Orthodontics. Orthodontists deal with with abnormalities in tooth and jaw relationships that result in facial distortion and malfunction. Orthodontic treatment attempts to establish normal closing of the mouth and facial harmony. The teeth are positioned and the jaws modified by fixed or removable appliances. Some habits such as tongue thrust must be corrected. The teeth are held in the new position during modification of the bone and muscles. Successful treatment results in normal shape and expression of mouth and lips, better enunciation, and good mastication. A major advance has been the use of agents that directly attach appliances to the teeth, in contrast to metal bands. The subspecialty surgical orthodontics is concerned with correcting the position and relationship of teeth and jaws by surgery.

Pedodontics. Pedodontics, the practice of dentistry in children, requires specialized knowledge and skills. The deciduous or primary teeth are small and differ in shape from adult teeth; special procedures are required for their treatment, which is made difficult by the immaturity of the patient. Caries are extremely prevalent in some groups of children, and premature tooth loss has serious effects on the permanent dentition. Harmful habits in children include thumbsucking, which displaces the teeth, and the consumption of large quantities of sugary foods, which causes decay. Caries and periodontal disease can be prevented by tooth cleaning, in which it is important to train children. In addition to the usual dental education, pedodontists are trained in behavioral science, hospital dentistry, anesthesiology, and human growth and development.

Periodontics. This branch of dentistry is devoted to diseases of the gingiva (gum tissue), alveolar bone, periodontal ligament, and cementum. Periodontal diseases include gingivitis, periodontitis (sometimes called pyorrhea), primary and secondary occlusal traumatism, gingival hyperplasia, and periodontal atrophy. The first two are the most important. Both are caused by bacteria at the junction between the tooth and the gingival tissue. Several types of anaerobic microorganisms are associated with periodontal destruction, including serotypes and species of *Actinobacillus*, *Bacteroides*, *Eikenella*, *Wolinella*, and *Fusobacterium*. Gingivitis is diagnosed when the inflammation is confined to the gingiva; in periodontitis the inflammation extends into the bone and periodontal ligament. In chronic periodontitis, pockets form between the tooth surface and the periodontal tissues, the periodontal ligament and alveolar bone are destroyed, an abscess may form, and the tooth can be lost.

The objectives of periodontal therapy are to eliminate the infection and to prevent its recurrence. Therapy consists initially of removal of microbial deposits (calculus and plaque) from the surfaces of the affected teeth by scaling and root planing with or without opening flaps, and adoption of meticulous oral hygiene. In cases of early-onset aggressive periodontitis, antibiotics may be used in addition to scaling and root planing. Nonsteroidal anti-inflammatory drugs may be useful adjuncts. In some cases, surgery may be needed to gain access to diseased tooth roots and eliminate periodontal pockets to prevent recurrent infection. Periodontics includes implantology. *See* PERIODONTAL DISEASE.

Prosthodontics. Prosthodontics is the branch of dentistry devoted to artificial substitutes for oral structures. The replacement of teeth and other oral structures may be necessitated by abnormalities, loss of teeth from disease or trauma at birth, and destruction of teeth or jaws or other parts of the mouth by surgical treatment of neoplasms or trauma. Prosthodontics has two subdivisions: removable prosthodontics, in which missing teeth are replaced by dentures; and fixed prosthodontics, in which teeth are replaced by bridges. Bridges serve when healthy teeth or implants are present at the ends of an edentulous (toothless) space. The support of bridges by implants is an important technique. In addition, prosthodontists use splints to connect teeth that have been loosened by periodontal disease. They also construct devices that modify the relationships of the biting surfaces of the teeth to alleviate facial and joint pain.

Oral pathology. Oral pathology is concerned with the diseases of the teeth, oral cavity, and jaws, and with the oral manifestations of systemic diseases. Oral pathologists correlate clinical and radiographic features with the history of the tissue changes associated with disease to reach a diagnosis and in some cases to suggest treatment. *See* PATHOLOGY.

Pubic health dentistry. This is the science of promoting dental health through community effort. It comprises research, education, prevention, diagnosis, prescription, and the evaluation of community dental care. Public health dentists must have skill in public and human relations, a knowledge of administrative practice, and the ability to determine the needs of the community for dental care facilities and to assist in their design, development, administration, and support. They must assist in the development of dental health education programs, and possess a broad understanding of the social and behavioral sciences as these affect dental disease in the community. In many communities they have sponsored the fluoridation of water systems for controlling caries. *See* MOUTH DISORDERS; TOOTH DISORDERS.

Roy C. Page

Dental Materials

Dental materials is an interdisciplinary area that applies biology, chemistry, and physics to the development, understanding, and evaluation of materials used in the practice of dentistry. It is principally involved in restorative dentistry, prosthodontics, pedodontics, and orthodontics, and to a smaller extent in the other areas of practice. The field of dental materials has advanced rapidly with the application of technologies such as genetic engineering of filling materials, computer designing and machining of restorations, adhesive bonding to teeth, dental implants, and new types of restorative materials.

Restorative materials. The restoration of missing tooth structure is one of the most challenging areas

for the use of dental materials. The enamel that covers the crown of the tooth is the hardest substance in the human body and is very resistant to physical and chemical attack. Direct restorations are put in place and harden by a setting reaction, whereas indirect restorations require a replica of the tooth which has to be fabricated.

Gold. Pure gold or gold foil has long been used as a restorative material. Pellets of pure gold are formed from very thin gold sheets or foils. The gold is compacted with compressed air or manually, and contoured and polished to the desired shape. These restorations are long-lasting and biologically inert. They are placed in low-stress areas since gold is soft.

Amalgam. Dental amalgam is the most frequently used direct restorative material for posterior (molar and premolar) teeth. It is formed by mixing mercury with an alloy powder of silver, tin, and copper. After mixing, there is time to insert and shape the material before it hardens. Silver–copper compounds and elements such as indium improve the alloy. Its advantages are ease of application, durability, and cost; the disadvantages are chiefly esthetic. *See* AMALGAM.

Resins. Composite resins are tooth-colored filling materials that were first developed for restoring anterior teeth (incisors and canines) but have found increased use as an attractive alternative to dental amalgam in posterior teeth. A composite resin consists of an organic polymer matrix reinforced with up to 80% of an inorganic filler such as silica or glass. Composite resin from a single paste is placed on the tooth and then polymerized with the help of a fiber-optic visible light source. This allows more time for contouring and gives a restoration with improved color stability. Composite resin restorations are strengthened by etching the enamel with 37% orthophosphoric acid prior to placement of the resin in order to create a strong attachment to the enamel.

Glass ionomer. The only dental restorative material that forms a durable chemical bond to dentin is formed by the reaction of aluminosilicate glass with polyacrylic acid. Glass ionomer is particularly useful for restoring eroded or carious areas on exposed root surfaces with little or no enamel for bonding with composite resin. This material slowly releases fluoride to provide some additional protection from new decay. It is brittle and not used in areas of heavy stress.

Metal castings. Crowns and inlays are made of metal castings and used when there is insufficient tooth structure to support filling. A negative of the tooth is prepared with synthetic rubber or agar, into which high-strength gypsum (plaster) is poured. This forms the replica of the tooth, to which wax is applied. An indirect lost-wax process is used in fabricating dental castings. Alloy ingots used for the castings cover a range of compositions from 80% of noble metals (gold and the platinum group) combined with copper and other minor elements to base metal alloys of chrome, cobalt, and nickel. Traditionally, only noble alloys were used, but as the cost of gold increased, other alloys were developed. For making the impression and attaching the casting and adjusting the bite, at least two sessions are required. *See* METAL CASTING.

Porcelain. Porcelain jacket crowns and metal crowns to which porcelain is bonded are used in visible areas. In one type an inner core of high-strength aluminous (40–50% alumina) porcelain is covered by feldspathic porcelain. The latter is applied to a refractory gypsum model covered with platinum foil and sintered under a vacuum at a high temperature. The porcelain powder contains metal oxides to color the crown. The other type is made from a glass ceramic of silica that is cast by the lost-wax process. After casting and contouring, the crown is heat-treated to form an opaque crystalline material, mica. Finally, color is applied by firing oxide stains on the surface. Porcelain and cast ceramics are used for a cosmetic veneer of the front surface of anterior teeth. The thin veneers are bonded to etched enamel with a composite resin cement. The third type of crown has a cast metal substructure for strength, onto which feldspathic porcelain is bonded by the same process described above. These crowns resist chewing most successfully, whereas the jacket crowns are considered more attractive because they resemble enamel in their translucency.

Provisional restoration. For provisional restorations different materials are frequently used. While a permanent crown is being fabricated, a temporary or provisional crown protects the tooth. For best appearance, a custom-formed poly(methyl methacrylate) or polycarbonate crown is lined with acrylic resin. Aluminum crowns lined with acrylic resin can be used with posterior teeth along with the temporary crowns. These crowns are easily removed, since a cement is used that sets when pastes containing the active ingredients zinc oxide and eugenol are combined. A zinc oxide–eugenol cement reinforced with poly(methyl methacrylate) beads is also used as a temporary direct filling material for patients with deep decay. *See* POLYACRYLONITRILE RESINS.

Prosthodontic dental materials. The materials used with fixed prosthodontics are similar to those described above under "Metal castings" and "Porcelain." For small areas where teeth have been lost, a fixed partial denture (bridge) restores the missing teeth. This typically consists of crowns on the teeth next to the toothless space, rigidly connected with a casting suspended over the gap. Given multiple toothless areas on the same arch or long spaces, a removable partial denture can be placed that consists of a thin chrome–cobalt framework that rests on existing teeth. Highly cross-linked poly (methyl methacrylate) resin, pigmented to match the gingival tissue, is attached to the framework in the toothless areas for additional support of the artificial teeth and for appearance. Plastic, composite resin, or porcelain denture teeth are bonded to the resin base to recreate the missing occlusion. In the case of a completely toothless area, a complete removable denture is inserted, consisting of the acrylic resin base and teeth mentioned above. There has been increasing use of carefully placed titanium implants into bone where teeth have been lost. An individual crown or a partial or complete denture is placed on the implants

projecting through the gingival tissue. The crown or denture is secured with screws so that the artificial teeth can be removed for cleaning, inspection, and maintenance.

Orthodontic materials. The movement and stabilization of teeth can be controlled by several unique dental materials. Stainless steel alloys are commonly used for brackets and bands on teeth, but plastic brackets are also being used to improve appearance. Brackets are bonded to etched enamel with the help of composite resin without inorganic filler. Bands encircling posterior teeth are made of stainless steel and are generally secured with a cement formed by combining zinc oxide and phosphoric acid. Wires in different configurations are adapted to the dental arch to create the desired static forces for controlled tooth movement. The wires are commonly formed from wrought stainless steel, but alloys of nickel–titanium, titanium–molybdenum, and cobalt–chromium–nickel are also used. At one time, thin stainless steel wires were used to secure the archwire to the brackets, but today rubber elastic bands are preferred. Removable appliances made of poly(methyl methacrylate) are frequently used to move or stabilize individual teeth, or expand the dental arch. Stainless steel wires protruding from the acrylic generate active or passive forces on the teeth.

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Dentition

The arrangement, type, and number of teeth which are variously located in the oral or the pharyngeal cavities, or in both. Teeth are found in areas where there is an underlying supporting structure of cartilage or bone and where stomodeal ectoderm is present. The bones with which these structures are usually associated are the mandible (both the cartilaginous bones and the dentary), the premaxillaries, and the maxillaries. However, in certain vertebrates, the vomerine, palatine, parasphenoid, and pterygoid bones may be involved; and in many fishes, teeth are also found on the branchial arches.

Attachment and replacement. Attachment of teeth is variable among vertebrates. They may be inserted in sockets in the jawbones (thecodont condition), fused to the edge of the bone proper (acrodont condition), or attached to the inner surface of the jawbone (pleurodont condition). In polyphyodont animals, teeth may be constantly replaced. The polyphyodont condition is characteristic of fishes and the lower tetrapods in which the teeth are replaced as they are worn out. Replacement may be random, as in many fishes (sharks are an excellent example), or may follow a definite pattern, such as the waves of replacement seen in many reptiles. Most mammals have two sets of teeth during their lifetime, which is the diphyodont condition. These sets are the deciduous, or milk, teeth and the permanent dentition. Monophyodont dentition is the development of only one set of teeth.

Dental formulas of some mammals

Animal	Teeth				Total
	I	C	Pm	M	
Human	2/2	1/1	2/2	3/3	32
Cony	3/3	1/1	4/4	4/4	48
Beaver	1/1	0/0	1/1	3/3	20
Cat	3/3	1/1	3/2	1/1	30
Dog	3/3	1/1	4/4	2/3	42
Sheep	0/3	0/1	3/3	3/3	32
Lynx	3/3	1/1	2/2	1/1	28
Rat	1/1	0/0	0/0	3/3	16
Horse	3/3	1/1	4/4	3/3	44
Mole	3/3	1/1	4/4	3/3	44
Squirrel	1/1	0/0	2/1	3/3	22
Reindeer	0/3	0/1	3/3	3/3	32
Pig	3/3	1/1	4/4	3/3	44
Common seal	3/2	1/1	4/4	1/1	34
Skunk	3/3	1/1	3/3	1/2	34
Raccoon	3/3	1/1	4/4	2/2	40
Bear	3/3	1/1	4/4	2/3	42

Patterns. Teeth may have a similar form in all regions where they occur. Such teeth are homodont, as distinct from those that are variable in shape, called heterodont dentition. The homodont condition is characteristic of nonmammalian vertebrates, although the anterior teeth may differ from those lying in the cheek region of the jaws. A heterodont dentition began to evolve in the reptilian ancestors of mammals, but became fully developed in the mammals. The mammalian heterodont condition is frequently described by a dental formula expressing both the number and kind of teeth in each half jaw, both upper and lower (see **table**). For humans the formula is I 2/2, C 1/1, Pm 2/2, M 3/3. This can be read for either the right or left side of the jaw; there are 2 upper and 2 lower incisors, 1 upper and 1 lower canine, and 2 premolars and 3 molars in both the upper and lower jaws. According to this formula, humans have a total of 32 teeth, 8 located in each half of the upper jaw and 8 in each half of the lower jaw.

The initial letter may be omitted in the dental formula, and if a tooth does not appear in the group, a zero so indicates. Unfortunately, the formula does not indicate which of the several molars or premolars may be absent in a species, nor does it provide information as to the morphology of the several tooth types. Such data are equally important. Information about the dentition is an essential part of mammalian systematics at all levels for both fossil and living species. The dental formula and other dentition information are included in the description of all species of living mammals. Because teeth are very hard structures and fossilize well, and because teeth are usually

characteristic of the mammalian taxa (even species in many cases), the dentition is one of the most valuable features in mammalian paleontology.

Evolution. Study of dentitions has revealed that the general evolutionary trend among vertebrates has been a reduction of the numerous successive dentitions in favor of fewer sets of teeth, and a change from the usual homodont dentition to the heterodont dentition of mammals with its several characteristic types of teeth. The latter trend has reversed in some mammalian groups, such as the whales, which have a very uniform homodont dentition. *See* TOOTH.

Frederick S. Szalay

Denudation

The term "denudation" was introduced into the earth sciences in the early nineteenth century to refer to the removal of surface material that leads to the exposure of underlying strata and rock units. Now, denudation refers to all the weathering and erosional processes that contribute to the lowering of the land surface, potentially to the production of a peneplain as the product of a long period of erosion. Denudation is, thus, the complement of deposition, the accumulation of the products of denudation in sedimentation basins. *See* BASIN; DEPOSITIONAL SYSTEMS AND ENVIRONMENTS; EROSION; WEATHERING PROCESSES.

Contemporary records of sediment and solute flux through river systems, representing the mass removed from the continental surfaces, have been used to estimate rates of denudation since the mid-nineteenth century. These are usually expressed as an average rate of lowering of the land surface, in units of millimeters per thousand years (mm/ka), and are often differentiated as mechanical denudation and chemical denudation. The former should include both suspended and bed load transport, but is often estimated from the suspended sediment yield alone, or with an added fixed proportion (10–15%) to account for bed load. The latter includes material removed from the landmasses as solutes within river discharges and usually needs to be corrected for those constituents which are derived from the atmosphere. *See* SEDIMENTOLOGY.

Even when estimated for large drainage basins, rates of denudation are highly variable. Mechanical denudation in the world's largest river systems varies between 4 mm/ka and more than 500 mm/ka. Chemical denudation rates on a global scale are both lower and less variable, falling within the range 2–30 mm/ka. This variability is greatly increased when denudation rates are calculated for small drainage basins, especially those in regions of variable relief. Environmental controls on contemporary rates of denudation have often been sought in climate, as reflected in effective precipitation, runoff, or vegetation. Other studies emphasize the importance of topography and local relief as controls of denudation rates, especially in tectonically active mountain systems.

On a geologic time scale, denudation should lead to the development of planation surfaces of low relief that were classically interpreted as providing a denudation chronology based on relative age. Until recently, such studies suffered from a dearth of procedures capable of defining a true chronology of erosional landforms. That has greatly changed in recent decades with the development of methods which allow the definition of denudation rates over geologic time scales. *See* GEOLOGIC TIME SCALE.

Attempts to invert sedimentary and stratigraphic records, and so define rates of erosion on the contributing continental surface, have greatly benefitted from improved seismic surveys, core extraction, and modeling of sedimentation basin processes. On the land surface itself, estimates of denudation rates based on the dissection of surfaces of known age have improved with the use of digital terrain models and absolute dating techniques. Equivalent estimates show that very rapid rates of bedrock denudation (up to 15 mm/yr) have been maintained in the western Himalaya during the past few million years. *See* DATING METHODS.

However, developments in the dating of bedrock surfaces and surficial material in the 1990s represent the greatest improvement in defining rates of denudation during the late Cenozoic era. Potassium-argon (K-Ar) dating of lavas and the magnitude of dissection on volcanic cones had been used earlier to define erosion rates. More recently, exposure dating of bedrock through the accumulation of cosmogenic nuclides produced in situ within them is a development that allows the estimation of average surface lowering rates over long time intervals. Cosmogenic nuclides have also been used to define rates of regolith production and erosion, contributing to denudation, over more recent time periods. Fission track analysis of apatite is another recent development that has had success in defining the long-term tectonic and denudational history of continental margins and mountain systems. *See* APATITE; CENOZOIC; COSMOGENIC NUCLIDE; FISSION TRACK DATING; REGOLITH.

Nel Caine

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Deoxyribonucleic acid (DNA)

A linear polymer made up of a specific sequence of deoxyribonucleotide repeating units linked by 3′,5′-phosphodiester bonds; it is the carrier of genetic information. The set of DNA molecules that contains all genetic information for an organism is called its genome. DNA is found primarily in the nuclei of eukaryotic cells and in the nucleoid of bacteria. Small amounts of DNA are also found in organelles (such

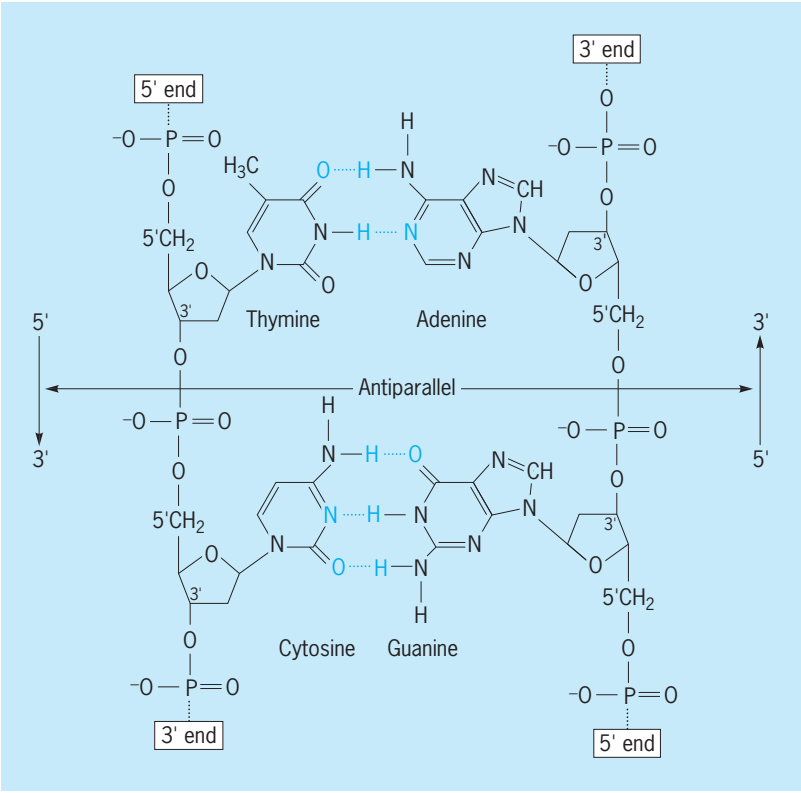

Fig. 1. Diagram of the DNA duplex. Each of the two strands comprises repeating units of a sugar–phosphate polymer. Each nucleotide on one strand is connected to the next by a phosphodiester bond. The linkage is always made from the the 3′ position of one sugar to the 5′ position of the next. The two strands are aligned antiparallel to each other, with deoxyribose on the outside and purines and pyrimidines on the inside of the duplex. Purines and pyrimidines share hydrogen bonds, with adenine binding to thymidine and cytosine binding to guanine. (*Reprinted with permission of John Wiley & Sons, Inc., from T. M. Devlin, Textbook of Biochemistry with Clinical Correlations, 4th ed., 2002. Copyright © 1997 Wiley-Liss, Inc.*)

as mitochondria and chloroplasts) that contain their own genomes, in autonomously maintained DNAs called plasmids, and in viruses. *See* NUCLEIC ACID.

DNA is composed of two long polymer strands of the sugar 2-deoxyribose, phosphate, and purine and pyrimidine bases. The backbone of each strand is composed of alternating 2-deoxyribose and phosphate linked together through phosphodiester bonds (**Fig. 1**). A DNA strand has directionality: each phosphate is linked to the 3′ position of the preceding deoxyribose and to the 5′ position of the following deoxyribose. The four bases found in DNA are adenine, thymine, guanine, and cytosine. Each 2-deoxyribose is linked to one of four bases in the 1′ position via a covalent glycosidic bond. A base linked to a sugar is called a nucleoside, and a base linked to a sugar phosphate is called a nucleotide. The numbering of atom positions of the sugar is accompanied by a prime sign to distinquish it from the positional numbering of the bases. The sequence of these four bases allows DNA to carry genetic information. Bases can form hydrogen bonds with each other. Adenine forms two bonds with thiamine, and cytosine forms three bonds with guanine. These two sets of base pairs have the same geometry, allowing DNA to maintain the same structure regardless of the specific sequence of base pairs (Fig. 1). *See* DEOXYRIBOSE; PURINE; PYRIMIDINE.

Structure. DNA can be single-stranded or aligned to form two-, three-, or even four-stranded structures. It is commonly found in its double-stranded form in chromosomes. In double-stranded DNA, the two strands wrap around each other to form a double helix (**Fig. 2**). The two strands are held together by base pairing and are antiparallel. Thus if one strand

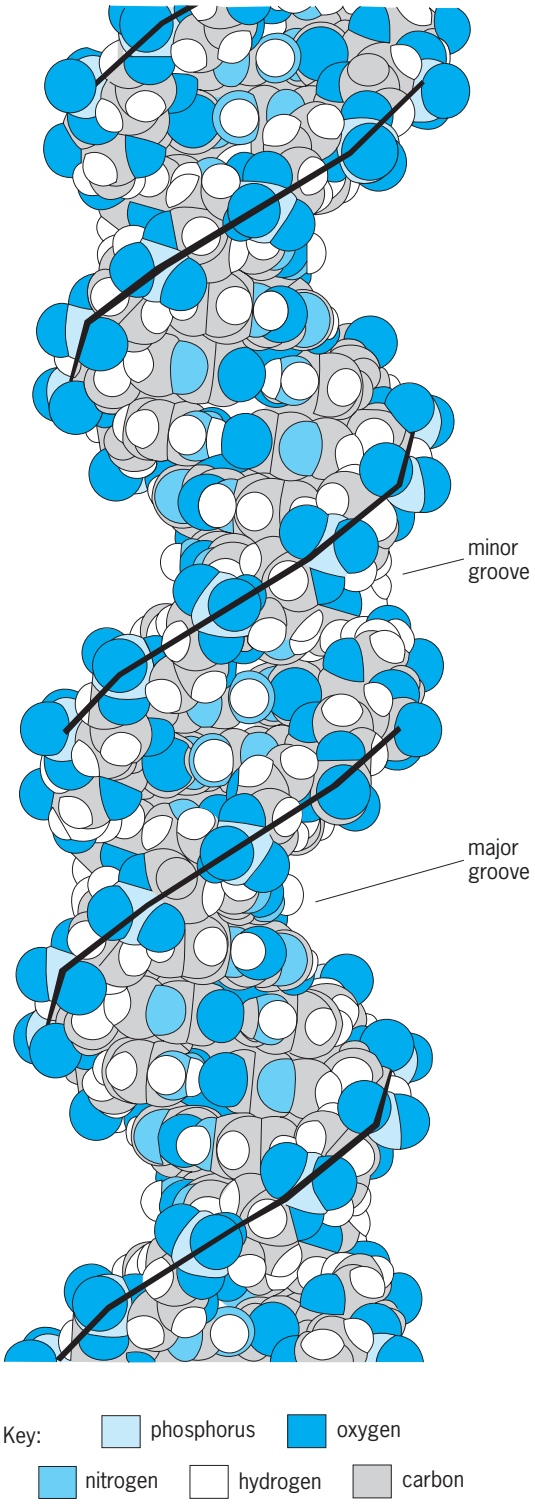

Fig. 2. B-DNA structure. The helical axis is vertical, and a solid line is drawn from phosphate to phosphate along the backbone.

is oriented in the 5′ to 3′ direction, the other strand will be 3′ to 5′. This double-helical structure of DNA was first proposed in 1954 by J. D. Watson and F. H. C. Crick.

The most common form of DNA is the B-form, which is a right-handed double helix with 10.4 base pairs per turn (Fig. 2). In a B-form double helix, the ribose–phosphate backbones run along the outside of the helix, and the base pairs are roughly perpendicular to the axis of the helix and lie just offset from the center axis of the helix. This offset causes one of the two grooves that run along the helix to be larger than the other—thus the designations, major groove and minor groove. Proteins that bind to DNA usually bind specifically to one of the two grooves. Less common forms of DNA include the A-form, which is a right-handed double helix that has 11 base pairs per turn and has a wider diameter than the B-form, and the Z-form, which is a narrow, irregular left-handed double helix (**Fig. 3**).

Right-handed conformation. Following Watson and Crick's proposal in 1954, experiments supported a right-handed form of the double helix in which there were approximately 10 base pairs per turn. In this form, each base pair is similar to the neighboring base pair, except that there is a rotation of approximately 36° between the base pairs. The conformation of DNA was determined by x-ray diffraction studies of fibers. A dried DNA fiber produces a type A pattern, in which the base pairs are tilted somewhat away from the helix axis. However, a wet or hydrated DNA fiber produces a B-type pattern, in which the bases are perpendicular to the helix axis (Fig. 2).

Left-handed conformation (Z-DNA). In 1979 it was discovered that DNA can also form a left-handed double helix with the same hydrogen bonding between the two bases (Fig. 1). This was discovered by solving the structure of crystals of short DNA fragments; crystals can yield a more detailed view of the molecule. A solid line drawn between the phosphates in the left-handed conformation follows a zigzag path (Fig. 3); hence it is called Z-DNA. In contrast to B-DNA, Z-DNA has only one deep helical groove extending almost to the axis of the helix. The base pairs in left-handed Z-DNA are located near the outside of the helix. B-DNA has a one-base-pair repeating unit, whereas Z-DNA has a dinucleotide repeating pattern. Both forms of DNA have the two sugar–phosphate backbones oriented in opposite directions (Fig. 1). Z-DNA has been identified in DNA fibers as well as inside living cells.

Conformational changes. In biological systems, DNA is normally subjected to a torsional strain that tends to unwind the double helix. This torsional strain tends to untwist right-handed B-DNA and sometimes is great enough to stabilize certain regions of DNA in the left-handed Z-DNA conformation. The conversion from right-handed B-DNA to left-handed Z-DNA is associated with changes in the relationship of the bases to the sugar backbone. Every other base in Z-DNA is rotated about the bond connecting it to the sugar. This altered form is common in adenine or guanine residues (purines) but less common

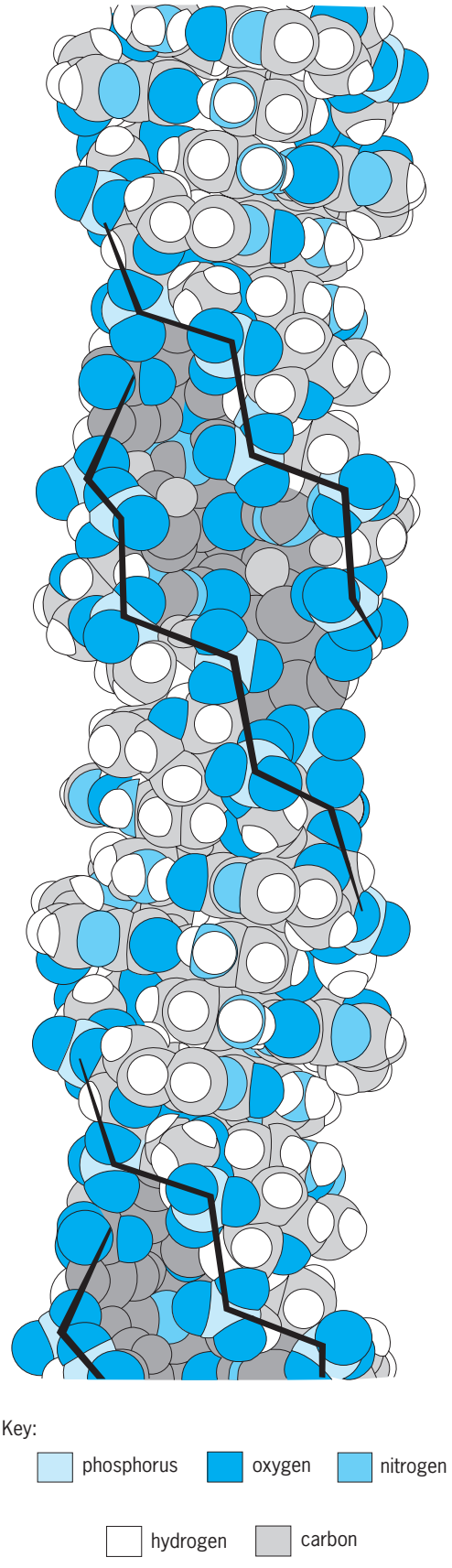

Fig. 3. Z-DNA structure. Darker gray area is the deep, narrow minor groove. A solid line is drawn from phosphate to phosphate along the backbone.

in thymine or cytosine residues (pyrimidines); thus, sections that form Z-DNA often have alternations of purines and pyrimidines.

Although right-handed B-DNA is the most common form of DNA, left-handed Z-DNA occurs in a number of places and may play an important role in the regulation of gene expression. In addition, Z-DNA appears to be an important component of the process of crossing-over or recombination of DNA molecules. Z-DNA is important in carrying out special functions that are in general associated with special proteins that bind to Z-DNA and not to B-DNA. *See* CROSSING-OVER (GENETICS); GENE ACTION.

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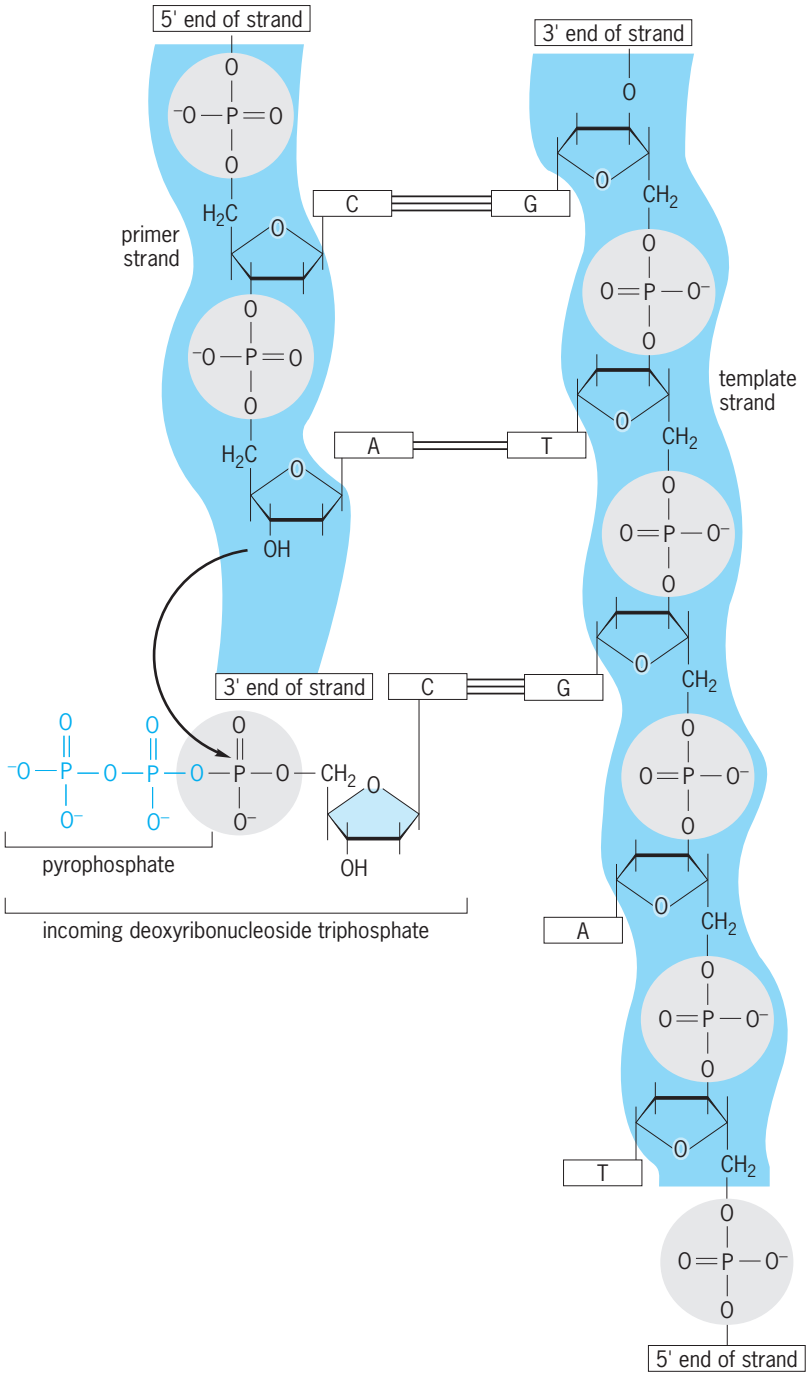

Fig. 4. Replication of DNA. The replication shown here highlights the selection of new base insertion by using Watson-Crick (A-to-T and G-to-C) base pairing, the addition of new bases to the 3′ end of an existing chain, and the attack of the 3′OH on the alpha phosphate of a nucleoside triphosphate to add the newest base. (*Copyright © 2002. From Molecular Biology of the Cell, 4th ed., by B. Alberts et al. Reproduced with permission of Garland Science/Taylor & Francis LLC*)

Function. For cells to live and grow, the genetic information in DNA must be (1) propagated and maintained from generation to generation and (2) expressed to synthesize the components of a cell. These two functions are carried out by the processes of DNA replication and transcription, respectively. *See* GENETIC CODE.

Replication. Each of the two strands of a DNA double helix contains all of the information necessary to make a new double-stranded molecule (**Fig. 4**). During replication the two parental strands are separated, and each is used as a template for the synthesis of a new strand of DNA. Base incorporation is directed by the existing DNA strand; nucleotides that base-pair with the template are added to the nascent DNA strand. The product of replication is two complete double-stranded helices, each of which contains all of the genetic information (has the identical base sequence) of the parental DNA. Each progeny double helix is composed of one parental and one nascent strand. DNA replication is very accurate. In bacteria the mutation rate is about 1 error per 1000 bacteria per generation, or about 1 error in 10^9 base pairs replicated. This low error rate is due to a combination of the high accuracy of the replication process and cellular pathways which repair misincorporated bases.

Synthesis of the nascent DNA strands is carried out by a family of enzymes called DNA polymerases. These enzymes synthesize a new phosphodiester bond between the 3′ hydroxyl position of the terminal deoxyribose of an existing DNA (or ribonucleic acid, RNA) chain and a deoxynucleotide triphosphate. All DNA polymerases (1) synthesize new DNA in a 5′ to 3′ direction, (2) require a primer (they must add on to the 3′ hydroxyl position of DNA or RNA), and (3) require a DNA molecule as a template. DNA polymerases, in conjunction with multiple accessory proteins, can rapidly synthesize new DNA strands. (The mandatory participation of the 3′ hydroxyl forms the rationale for the use of nucleoside analogues that lack a 3′ hydroxyl as antiviral and anticancer drugs, as they lead to premature chain termination.) A bacterial replication fork moves at a rate of approximately 1000 nucleotides per second, while in eukaryotic cells forks move at a rate of about 100 nucleotides per second. *See* MUTATION.

Transcription. In transcription, DNA acts as a template directing the synthesis of RNA. RNA is a single-stranded polymer similar to DNA except that it contains the sugar ribose instead of 2-deoxyribose, and the base uracil instead of thymidine (**Fig. 5**). The two strands of DNA separate transiently, and one of the two single-stranded regions is used as a template to direct the synthesis of an RNA strand. As in DNA replication, base pairing between the incoming ribonucleotide and the template strand determines the sequence of bases incorporated into the nascent

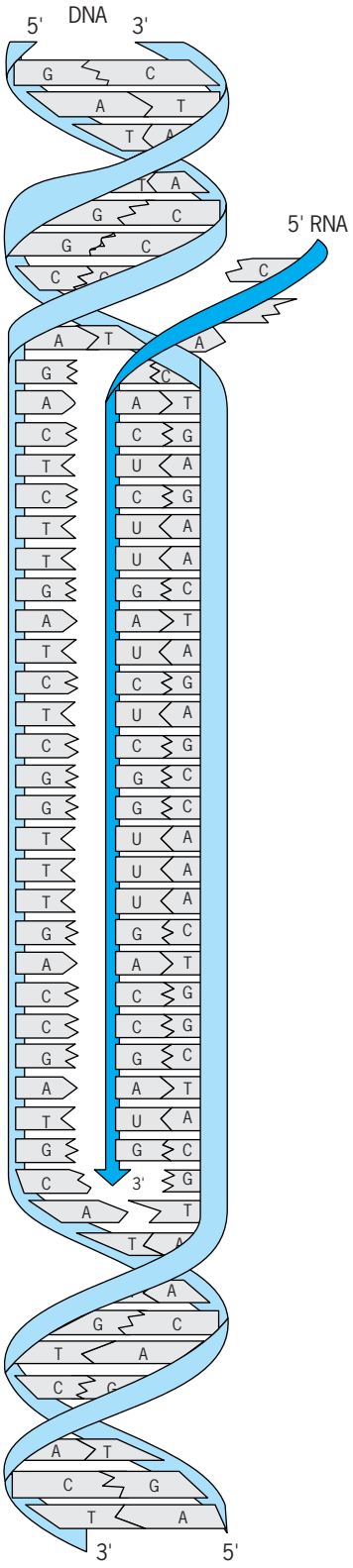

Fig. 5. Transcription of DNA. The double-stranded DNA is transiently separated to allow the synthesis of RNA. Using a mechanism similar to that discussed for DNA replication (Fig. 4), ribonucleotides are added to the 3′ end of a growing chain, and the selection of bases to insert is guided by Watson-Crick base pairing. (*Reprinted with permission from G. M. Cooper, The Cell: A Molecular Approach, 2d ed., 2000. © Sinauer Associates, Inc.*)

RNA. Thus, genetic information in the form of a specific sequence of bases is directly transferred from DNA to RNA in transcription. After the RNA is synthesized, the DNA reverts to a double-stranded form. Transcription is carried out by a family of enzymes called RNA polymerases. In prokaryotes, RNA polymerase interacts directly with the sites at which RNA transcription start (promoters), while in eukaryotes large complexes of proteins bind to promoters and then direct RNA polymerase to start RNA synthesis. Newly synthesized RNA is ultimately used to direct protein synthesis by ribosomes in a process called translation. *See* PROTEIN; RIBONUCLEIC ACID (RNA); RIBOSOMES.

Genetic variation. There is a great deal of variation in the DNA content and sequences in different organisms (**Fig. 6**). Because of base pairing, the ratios of adenine to thiamine and cytosine to guanine are always the same. However, the ratio of adenine and thymine to guanine and cytosine in different organisms ranges from 1:4 to 3:4. There is also large variation in the amount of DNA in the genome of various organisms. The simplest viruses have genomes of only a few thousand base pairs, while complex eukaryotic organisms have genomes of billions of base pairs. (The lungfish has one of the largest genomes known, about 100 billion base pairs.) This variation partially reflects the increasing number of genes necessary to encode more complex organisms, but mainly reflects an increase in the amount of DNA that does not encode proteins. A large percentage of the DNA in multicellular eukaryotes is noncoding or is repetitive DNA (sequences that are repeated many times). This DNA can serve structural or regulatory functions, it can be related to transposable elements (DNA sequences that can insert into other unrelated sequences), or in many cases it has no known function. Rarely, genes are present in multiple copies to allow high levels of expression of certain cellular components. For example, most cells have hundreds or thousands of copies of the genes encoding the components needed for protein synthesis to ensure sufficient levels of expression of these components.

In most eukaryotes the DNA sequences that encode proteins are not continuous but have other sequences interspersed within them. The sequences that are expressed in protein are called exons, and the interspersed sequences are called introns. The initial transcript synthesized by RNA polymerase contains both exons and introns and can be many times the length of the actual coding sequence. The RNA is then processed, and the introns are removed through a mechanism called RNA splicing to yield messenger RNA (mRNA), which is translated to make protein.

Recombinant technology. Techniques have been developed to allow DNA to be manipulated in the laboratory. These techniques, collectively called molecular biology or molecular genetics, have led to a revolution in biotechnology. This revolution began when methods were developed to cleave DNA at specific sequences and to join pieces of DNA together. Another major component of this technology is the

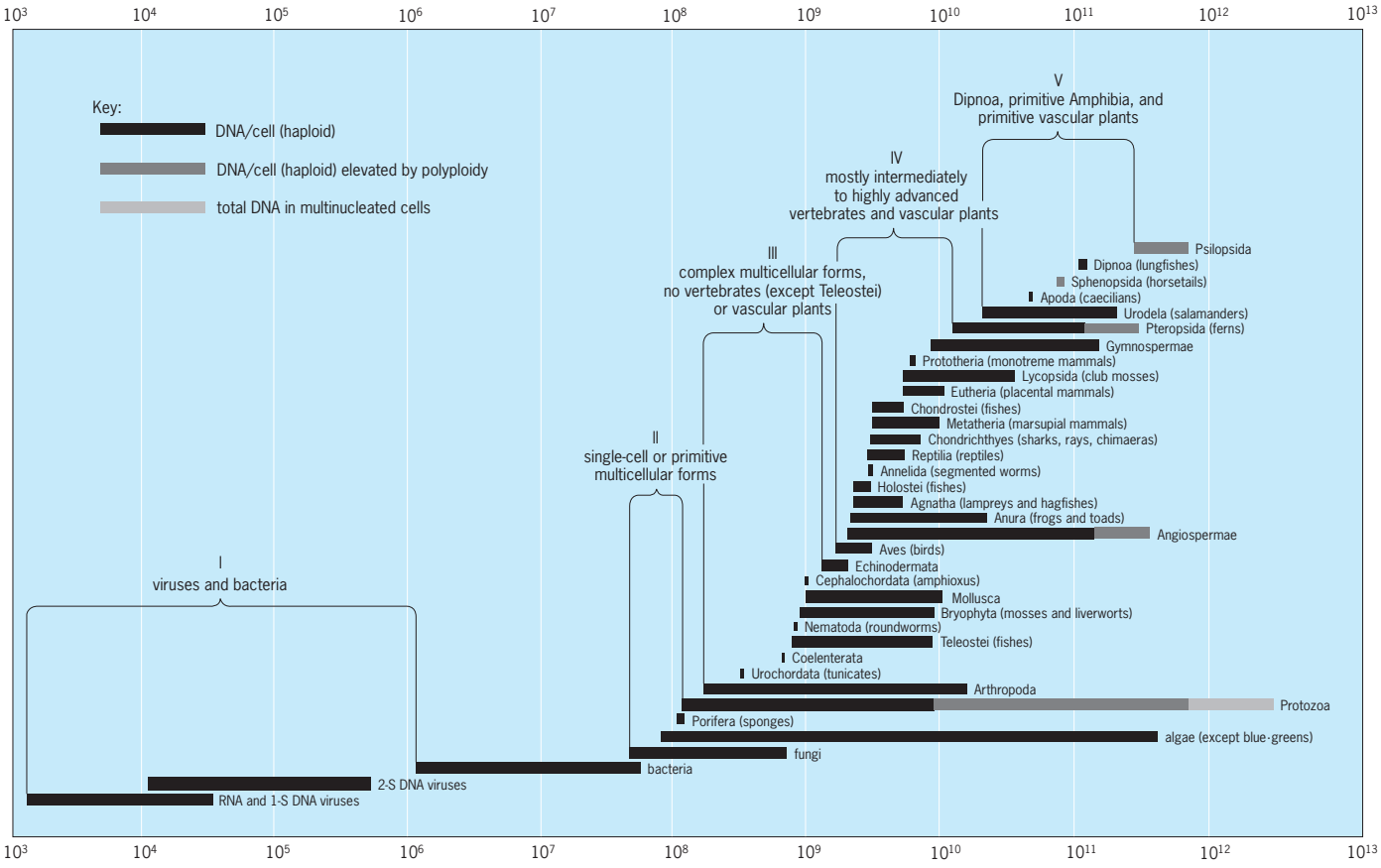

Fig. 6. Haploid DNA amounts (RNA in some viruses). Amounts are expressed as numbers of nucleotides. One gram of DNA = 2.01×10^{21} nucleotides. (*After A. J. Sparrow, H. J. Price, and A. G. Underbrink, A survey of DNA content per cell and per chromosome of prokaryotic and eukaryotic organisms: Some evolutionary considerations, in the Brookhaven Symposia in Biology, no. 23, Evolution of Genetic Systems, pp. 451-494, 1972*)

ability to determine the sequence of the bases in DNA. There are two general approaches for determining DNA sequence: either chemical reactions are carried out which specifically cleave the sugar–phosphate bond at sites which contain a certain base, or DNA is synthesized in the presence of modified bases that cause termination of DNA synthesis after the incorporation of a certain base. A series of four reactions, one for each base, are carried out, and the lengths of the DNA fragments produced identify the sequence of the bases. These methods can now be automated so that it is practical to determine the DNA sequences of the entire genome of an organism. The complete sequences of humans and a number of experimentally important organisms have been determined. *See* HUMAN GENOME PROJECT.

Another example of recombinant DNA technology is copy DNAs (cDNAs), which are made when an mRNA is converted to DNA using reverse transcriptase (a specialized DNA polymerase that can use RNA as a template). The resulting DNA can then be cloned, made in large quantity, and its sequence determined. Because it is much easier to manipulate DNA than RNA, cDNAs allow any RNA-based species to be characterized. *See* GENETIC ENGINEERING; MOLECULAR BIOLOGY; REVERSE TRANSCRIPTASE.

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In the cell. The full genome of DNA must be substantially compacted to fit into a cell. For example, virtually every cell in the human body contains a full genome of DNA, which in turn has a total length of about 3 m (10 ft). However, this DNA must fit into a nucleus with a diameter of 10^{-5} m. This immense reduction in length is accomplished in eukaryotes via multiple levels of compaction in a nucleoprotein structure termed chromatin. To summarize, the first level involves spooling about 200 base pairs of DNA onto a complex of basic proteins called histones to form a nucleosome. Nucleosomes are connected like beads on a string (**Fig. 7**) to form a 10-nanometer-diameter fiber, and this is further coiled to form a 30-nm fiber. The 30-nm fibers are further coiled and organized into loops formed by periodic attachments to a protein scaffold. This scaffold organizes the complex into the shape of the metaphase chromosome seen at mitosis.

The nucleosome is the fundamental structural unit of DNA in all eukaryotes. Virtually all of the DNA in each cell of an organism is packaged into nucleosomes. Nucleosomes are squat cylinders (6 nm × 11 nm × 11 nm) arranged as repeating subunits along the DNA at approximately 200-base-pair intervals. Pairs of four different histone proteins (H2A, H2B, H3, and H4) constitute an octamer complex,

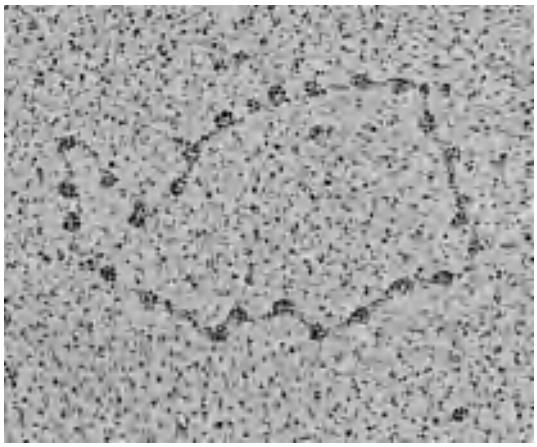

Fig. 7. Electron micrograph of the minichromosome of a virus that infects monkey cells. The circular viral DNA is organized in nucleosome "beads." In order to improve the ability to visualize individual nucleosomes, the conditions of microscopy in this figure have been adjusted to stretch the minichromosome slightly. In the cell, successive nucleosomes are in direct contact with one another, resulting in a circle of a relatively uniform fiber of about 11 nm diameter. (*From L. Lutter et al., Mol. Cell. Biol., 12:5004–5014, 1992*)

on the outside of which is wrapped 146 base pairs of DNA to form the nucleosome core. The DNA in the core is wrapped in $1^{3}/_{4}$ left-handed coils, resulting in approximately sixfold compaction of the DNA. The detailed structure of the nucleosome core (**Fig. 8**) has been determined by x-ray crystallography. Nucleosome cores are connected to one another by about 55 base pairs of linker DNA. The 146-base-pair length of DNA in the nucleosome core is invariant among all species studied, but the length of linker DNA can range from 20 to 100 base pairs among different species and different tissues. The reason for this variation is unknown. The linker DNA binds a fifth histone, H1, that plays a role in condensation of the 10-nm fiber into the higher-order structures.

Nucleosomes reduce the accessibility of DNA to DNA-binding proteins such as polymerases and other protein factors essential for transcription and replication. One level of regulation of gene expression occurs via posttranslational modification of histones. For example, specific protein factors activate transcription by adding acetyl groups to specific amino acid residues in the histones, while other factors with the opposite transcription effect remove those residues. Malfunctions in these regulatory processes have been associated with human diseases. There are several other types of histone posttranslational modifications with regulatory effects. Thus chromatin structure not only compacts the genome but also affords an additional level of control of gene expression. *See* CHROMOSOME; NUCLEOPROTEIN; NUCLEOSOME.

Leonard C. Lutter

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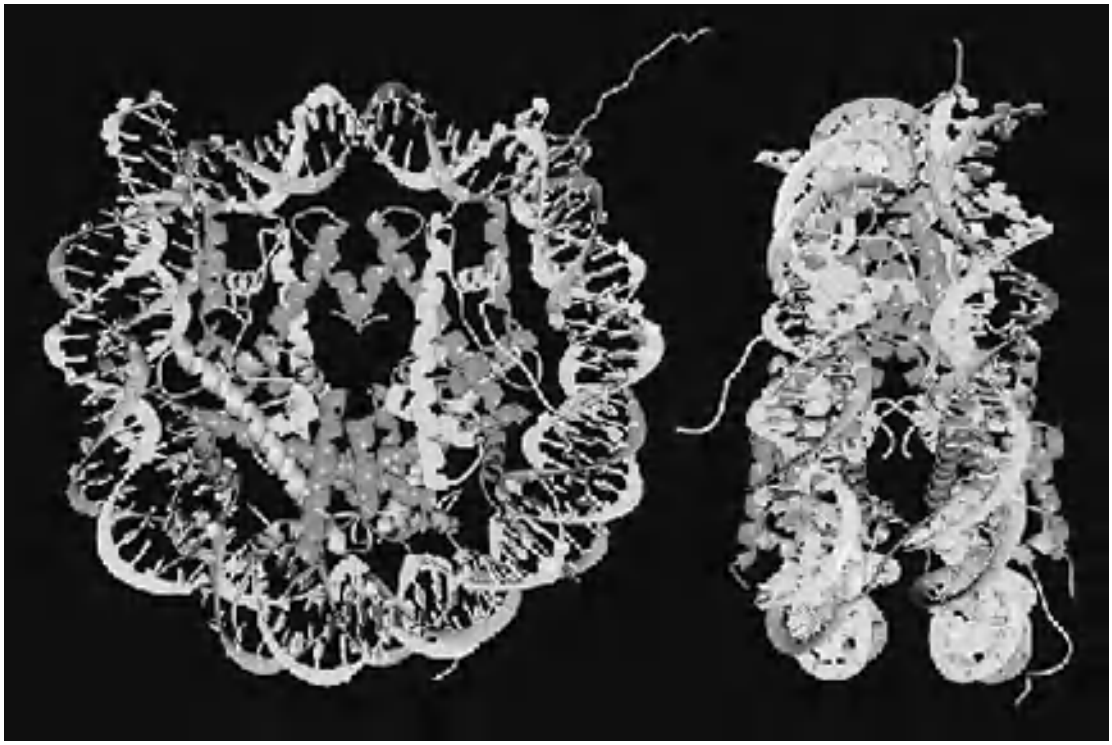

Fig. 8. Structure of the nucleosome core particle as determined by x-ray crystallography. On the left is a view down the superhelix axis, showing the DNA wrapped on the outside of the histone octamer. On the right is a view from the side, showing the two gyres of superhelix of 146 base pairs of DNA. (*From K. Luger et al., Nature, 389:251-260, 1997*)

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Depositional systems and environments

Depositional systems are descriptions of the interrelationships of form and the physical, chemical, or biological processes involved in the development of stratigraphic sequences. Depositional environments are the locations where accumulations of sediment have been deposited by either mechanical or chemical processes.

Depositional Systems

A basic premise in geological investigations is the concept of uniformitarianism—the present is the key to the past. This historical premise has been extended thoroughly in the fields of stratigraphy, sedimentology, and basin analysis to the point where ancient strata are commonly characterized by their genesis. Analysis of modern depositional systems and environments has long been the subject of geologic investigations, with studies increasing vastly in scope and volume since 1950. A greatly increased understanding of modern environments of deposition has generated a corresponding, extensive application of modern models to ancient strata that are inferred to be the analogs of modern depositional environments. *See* STRATIGRAPHY.

Traditional stratigraphic analysis, which emphasized the physically descriptive aspects, has thus changed; a critical genetic dimension has been added. Where once, for example, a formation may have been described in physical terms as a fine- to medium-grained quartzose sandstone, overlying a thick dark-gray shale sequence and underlying a coal-bearing sequence with discontinuous sands, the same sandstone may now be recognized as a delta-front sandstone, the underlying shale as a prodelta facies, and the overlying coal-bearing formation as the product of deposition on a delta plain. In this interpretation, the three distinct stratigraphic units become part of a genetically related sequence, each a component facies of a prograding delta system.

Modern systems. In a modern setting, a particular system of deposition is directly observable and known, whether it is a major delta, an alluvial fan, a meandering river, a barrier bar, a carbonate platform, or the like. Through specific observation, description, and delineation, it may be determined that a variety of depositional processes are active. At the terminus of rivers, sands may be deposited as bars in river mouths, creating the delta front; muds may be carried by suspension into the oceanic waters and deposited through flocculation, creating the prodelta; on the delta plain, a distributary channel may be depositing bed-load sands and, during flooding, may be carrying suspended muds to flanking flood basins or breaching the natural levee to

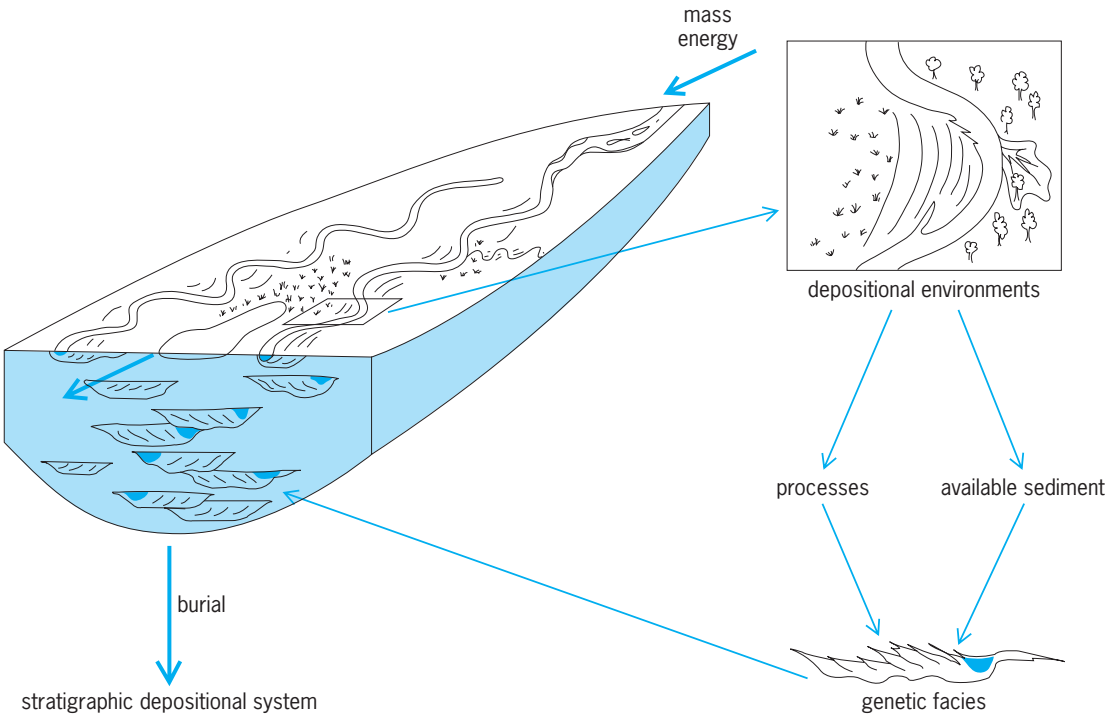


Fig. 1. An active depositional system consisting of genetically related environments. (*After W. E. Galloway, C. W. Kreitler, and J. H. McGowen, Depositional and Ground-Water Flow Systems in the Exploration for Uranium, University of Texas at Austin, Bureau of Economic Geology Special Publication, 1979*)

form crevasse splays. Each process or combination of processes gives rise to distinct, specific environments of deposition, with each resulting in a deposit which can be characterized by such features as lithologic composition, texture, sedimentary structures, geometry, size, and relationship to other deposits. *See* DELTA; FLOODPLAIN; RIVER.

Distinct physical, or in some cases biologic or chemical, products of deposition can be related directly to definable, operative processes. Such data permit the development of models of modern deposition in which processes and resulting deposits or facies are linked. By recognizing comparable physical, chemical, and biologic attributes of ancient strata, modern depositional analogs can be applied, and the original processes forming the ancient deposit can be inferred. Such an ancient deposit is called a genetic facies (**Fig. 1**). It contains the sedimentary record and constitutes a three-dimensional stratigraphic (ancient) depositional system. *See* FACIES (GEOLOGY).

Ancient record. As modern or active depositional systems consist of a complex of genetically related environments, the ancient sedimentary record likewise consists of a variety of genetic facies. When ancient genetic facies are defined and when several facies constitute a genetically related assemblage, they make up a depositional system (**Fig. 2**). Such systems, considered as three-dimensional entities, constitute specific units of entire sedimentary basin fills. A natural system may be defined as a structured set of objects, or attributes, which display recognizable interrelationships and operate together as a complex whole. The interdependence of parts is an inherent and diagnostic property of all natural systems.

An ancient depositional system is a three-dimensional, genetically defined, physical stratigraphic unit that consists of a contiguous set of process-related sedimentary facies. Several corollaries have evolved from the application of this concept. Depositional systems, such as delta, fluvial, and shelf systems, (1) are the stratigraphic equivalents of major physical geographic units; (2) form the principal building blocks of the sedimentary basin fill; and (3) can be applied where principal boundaries of the systems are preserved and where the geometry of the framework facies can be mapped.

Process-response systems. Depositional systems are process-response systems; that is, they describe the interrelationships of form and physical, chemical, or biologic processes. Such a process-response system provides a tool for prediction because it allows extrapolation beyond the data and can be used to predict properties not included in the original definition of the system. For example, the recognition of

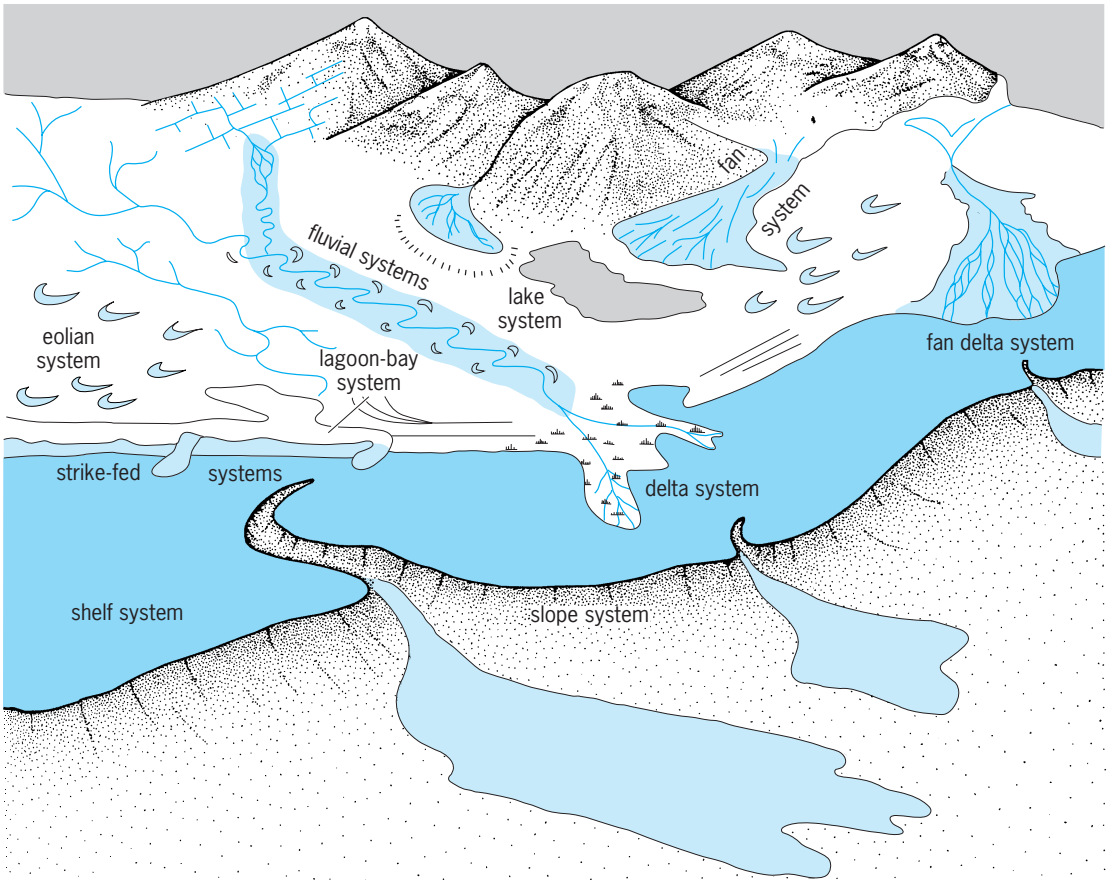


Fig. 2. Schematic setting of major depositional systems. (*After L. F. Brown, Jr., A. W. Cleaves III, and A. W. Erxleben, Pennsylvania Depositional Systems in North-Central Texas, University of Texas at Austin, Bureau of Economic Geology Guidebook 14, 1973*)

a delta-plain facies permits the prediction of the basinward occurrence of delta-front and prodelta facies, the logical component facies of a delta system.

The genetic facies and depositional system approach to analyzing ancient deposits is not a substitute for the traditional approaches to basin analysis. Rather, it utilizes all the traditional tools of analysis, particularly from the standpoint of description, but adds, through an analogy with modern processes and models, the genetic dimension.

Realms of deposition. The major realms of deposition may be classed broadly as terrigenous clastic depositional systems and biogenic-chemical depositional systems (**Figs. 3** and **4**). Each of these major systems is subdivided according to particular systems of deposition, and within each of the

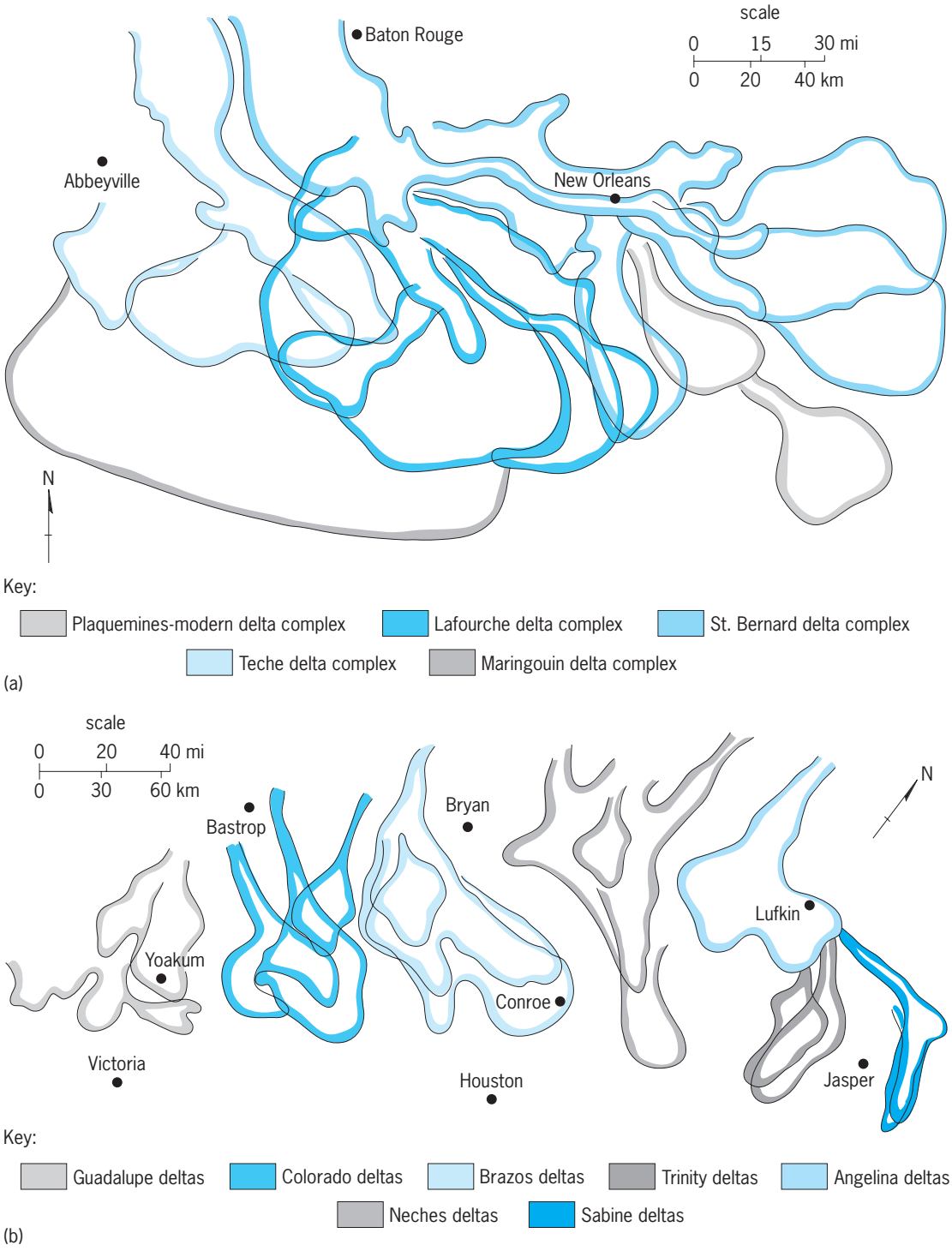
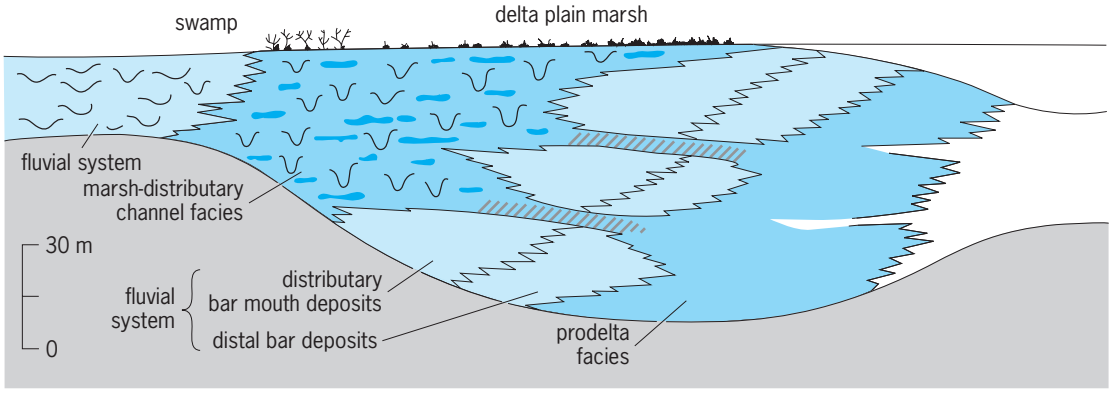


Fig. 3. Comparison of size, distribution, and arrangement of principal delta lobes of (*a*) the Holocene Mississippi delta system and (*b*) the Eocene Rockdale delta system of Texas. (*After W. L. Fisher and J. H. McGowen, Depositional systems in the Wilcox group of Texas and their relationship to the occurrence of oil and gas, Gulf Coast Ass. Geol. Soc. Trans., 17:105–125, 1967*)



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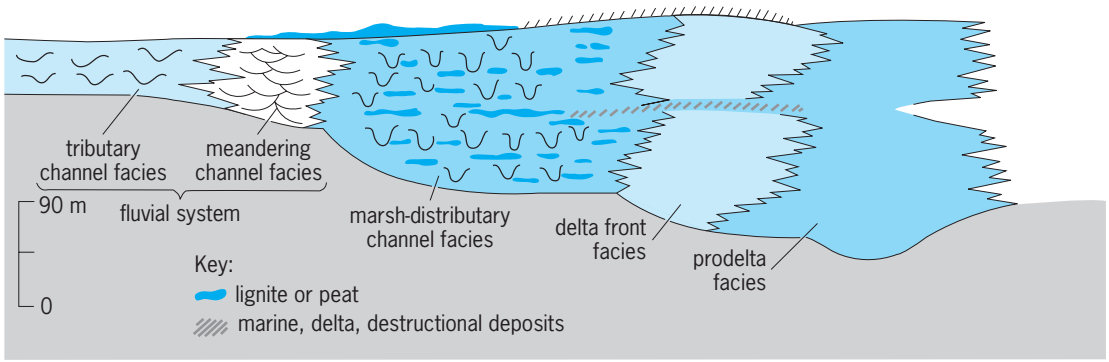

(b)

Fig. 4. Comparison of component facies of (*a*) the Holocene Mississippi delta system and (*b*) the Rockdale delta system of Texas. 1 m = 3.3 ft. (*After W. L. Fisher and J. H. McGowen, Depositional systems in the Wilcox group of Texas and their relationship to the occurrence of oil and gas, Gulf Coast Ass. Geol. Soc. Trans, 17:105–125, 1967*)

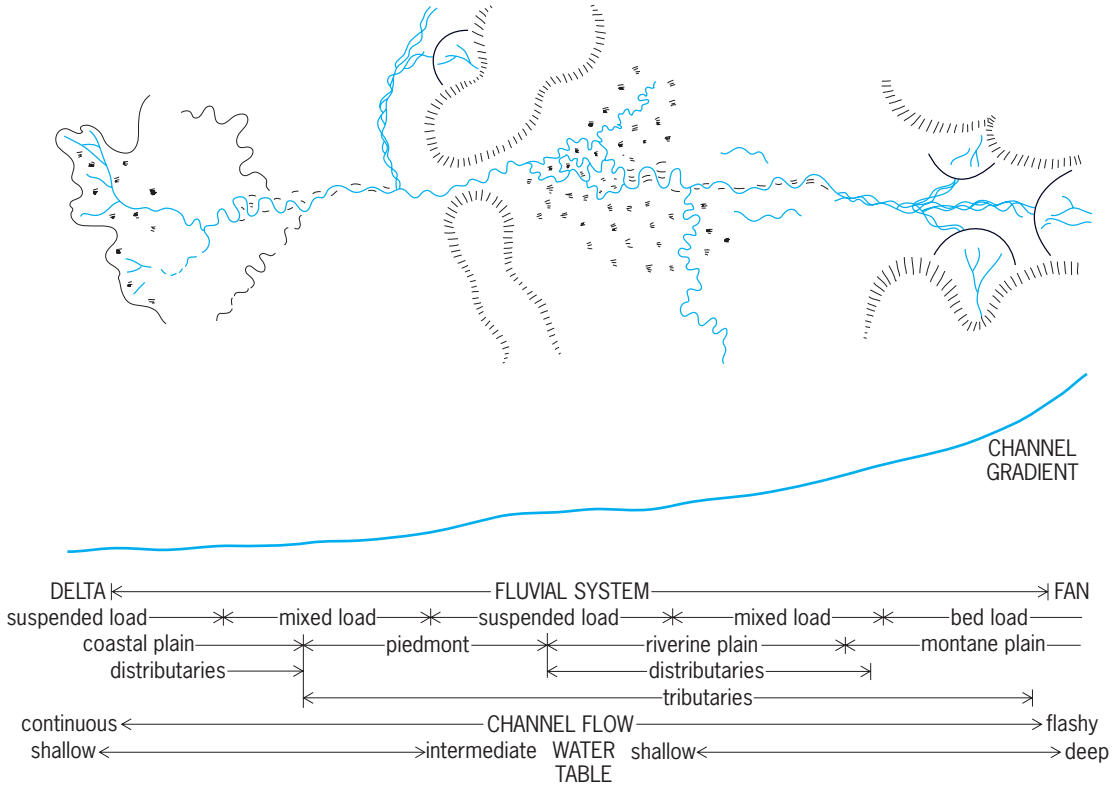

Fig. 5. Hypothetical fluvial systems. (*After W. E. Galloway, C. W. Kreitler, and J. H. McGowen, Depositional and Ground-Water Flow Systems in the Exploration for Uranium, University of Texas at Austin, Bureau of Economic Geology Special Publication, 1979*)

subdivisions is an assemblage of genetic facies, which are the fundamental units of depositional systems.

Terrigenous clastic systems, composed chiefly of sands and shales, embrace eight major systems: (1) fluvial or river systems, with braided stream systems, fine- and coarse-grained meandering stream systems, valley-fill stream systems, and straight distributary stream systems; (2) delta systems, with lobate and elongate river-dominated systems; (3) strike coastal systems, with barrier bar systems, strand-plain systems, bay and lagoon systems, and tidal flat systems; (4) fan or clastic wedge systems, with humid area and arid area alluvial fan systems and fan delta systems; (5) lacustrine systems, with open lake systems and closed or stratified lake systems; (6) continental eolian systems; (7) shelf systems, with wave- and tide-dominated systems; and (8) slope and abyssal systems, with tectonic up-lap slope systems, progradational off-lap slope systems, and destructive on-lap slope systems.

The biogenic-chemical systems consist of three major systems: (1) carbonate systems, with open and restricted carbonate platform systems, reef, bank, and atoll systems, winnowed shelf-edge systems, deep shelf systems, tidal flat and lagoonal systems, and slope systems; (2) glauconitic and authigenic shelf systems; and (3) evaporite systems, with supratidal (sebkha) systems and shallow shelf or platform and deep-water subaqueous systems.

By way of example, a hypothetical fluvial system may be considered (**Fig. 5**). A trunk channel or river evolves through bed-load (braided), mixed-load (meandering), and suspended-load phases in response to variations in gradient, sediment load, and discharge within the system. These phases, with resulting deposited genetic facies, may be characterized by differing attributes, recognizable or inferred, in both modern and ancient systems. Comparable kinds of variations, giving rise to distinct genetic facies, are recognizable in all major depositional systems.

W. L. Fisher

Depositional Environments

Depositional environments may be distinguished from erosional environments, in which erosion of the Earth's surface is taking place. Both depositional and erosional environments are of interest to geomorphologists. However, most attention to depositional environments has come from sedimentologists, particularly in order to understand the origin of sedimentary rocks. *See* EROSION.

Geophysical controls. The principal geophysical factors that control depositional environments are tectonics, climate, and topography.

Tectonics. Sediment is derived mainly from source areas that are actively undergoing uplift and erosion, and is deposited mainly in areas that are undergoing subsidence. Location of the source and basin of deposition is mainly controlled by large-scale geophysical processes acting within the Earth's mantle, so a major factor affecting the nature and distribution of sedimentary environments is the overall structural development, or tectonics, of the area. *See* PLATE TECTONICS.

Tectonics determines the major geological structure or setting of an environment of deposition, including the location and nature of the main areas undergoing uplift or subsidence. Areas with high relief, such as mountains and volcanoes, suffer rapid erosion and supply much more sediment to basins of deposition than larger areas of low relief. One investigation, for example, found that 82% of the suspended solids (mud) discharged by the Amazon River were supplied by the 12% of the drainage basin located within the Andes Mountains.

The major topographic slopes of continental surfaces are also largely determined by tectonic factors and tend to persist for long periods of geological time. The Mississippi River alluvial plain is an example. Large rivers have flowed south along the position of the modern Mississippi River plain for at least the last 200,000,000 years, and there have been many ancient precursors of the modern Mississippi delta and northern Gulf of Mexico coastal plain.

Tectonics is also one of the major factors controlling the rate of sedimentation in sedimentary environments. Local, short-term rates of accumulation of sediment may be very high. The modern Mississippi River delivers more than 1,100,000 tons (1,000,000 metric tons) of mud and sand to the delta every day; thicknesses of several meters of sediment may accumulate locally over periods of a few months. In the long term, however, the thickness of sediment that can accumulate in a shallow-water environment is determined by the regional rate of subsidence, which is determined by tectonic factors. A maximum of about 11 mi (18 km) of mainly shallow-water sediments has accumulated in the Mississippi region since the beginning of the Cenozoic, some 65,000,000 years ago. The long-term rate of sedimentation, therefore, has been only about 0.8 in. (20 mm) per year, which is still a very high rate of sedimentation averaged over long periods of geological time compared with most other areas.

The fact that the local, short-term supply of sediment is generally larger than the long-term rate of subsidence leads to rapid shifting of the centers of sedimentary deposition and therefore of sedimentary environments. In the Mississippi region, the position of the delta itself has switched abruptly several times in the last 5000 years. Sediment tends to fill in topographically low areas, not only in deltas but also in river and coastal plains and other environments supplied with large quantities of sediment. In one local area, rapid filling of a topographic low may be followed by a long period of nondeposition, when sediment is being deposited elsewhere. Sedimentation in the area starts again when subsidence (or a rise in sea level) once more produces a favorable environment for deposition. A thick sequence of sedimentary rocks, therefore, is frequently found to be composed of parts deposited rapidly during depositional episodes, separated by bounding surfaces or thin sedimentary units resulting from nondeposition

or very slow deposition over long periods of geological time.

Climate. A second important major control is climate. This includes the average temperature, the range of temperature variation, the aridity or humidity (ratio of evaporation to precipitation), and the magnitude and frequency of floods and storms. Climate in turn has an important influence on such physical factors as the salinity and energy of the environment (wind and water speeds and degree of turbulence, for example), as well as on the abundance and types of plants and animals.

Topography. In areas of subsidence and sedimentation, topography results from and controls sedimentary environments. Along a coastline of low relief, for example, spits and barrier islands are produced by waves generated in the open sea. Shallow lagoons on the landward side of barrier islands are protected from wave action by the islands themselves. The distinctive features of the lagoon environment are a result of a topography which has been produced by the accumulation of sediment in another sedimentary environment (the barrier island). *See* BARRIER ISLANDS.

Classification. Sedimentary environments can be classified into three categories: terrestrial, including alluvial fans, fluvial plains, sandy deserts, lakes, and glacial regions; mixed (shore-related), including deltas, estuaries, barrier island complexes, and glacial marine environments; and marine, including terrigenous shelves or shallow seas, carbonate shelves or platforms, continental slopes, continental rises, basin plains, ocean ridges, and ocean trenches.

Although the importance of tectonics and climate in controlling sedimentary environments is widely recognized, most classifications are based mainly on topography. Almost all distinguish terrestrial (subaerial or fresh-water) from marine environments, and also recognize an important group of mixed or shore-related environments. The above classification is just an outline: sedimentary environments are subject to almost endless subdivision, and many more complex classifications have been proposed.

Environmental controls. A number of major processes operate within environments and determine the types of sediment deposited in the environment.

Water depth. An important control is water depth, particularly in lakes and seas. Sediment deposited in shallow water is subject to constant agitation by wave action unless waves are excluded by some physical barrier. At extremely shallow depths, the sediment may be periodically exposed by changes in water level produced by tides, weather, or climatic fluctuations. Light easily penetrates to the bottom in very shallow water (unless the water is extremely turbid) so that aquatic plants, as well as animals, may flourish. Light is important not only for plants but for some organisms that live in close association with them. An example is hermatypic corals whose life cycle is so closely related to algae that typical coral reefs can develop only in water shallower than a few tens of meters. Criteria for shallow water depth in sedimentary deposits therefore include structures, such as ripple marks and cross-bedding, formed by strong currents, and particularly structures formed by wave action; structures produced by periodic exposure to the atmosphere, such as mud cracks; fossils or trace fossils (trails, burrows, and so on) indicating intense biological activity, and particularly algal growth (stromatolites, for example) and organically produced buildups of calcium carbonate (reefs and similar phenomena). Relatively deep water may be inferred from the absence of such phenomena, though there may be close resemblances between sediments formed in protected, quiet, shallow water and those deposited in deep water. *See* REEF; STROMATOLITE.

Energy of the environment. The degree of physical agitation by waves and currents, known as the energy of the environment, is frequently closely related to water depth. High-energy sedimentary environments include those in which the bottom sediment is constantly in motion, such as eolian sand dunes, the bottoms of river channels, beaches, and shallow marine environments characterized by strong waves or tidal currents. Fine sediment is maintained in suspension by the high level of turbulence in such environments and cannot accumulate unless there is an extremely high rate of supply. Typically the bottom sediment consists of well-sorted sand or gravel, with only a sparse growth of flora or fauna. Most organisms cannot flourish on a constantly moving sediment substrate. *See* DUNE; RIVER; TIDE.

Temperature. The temperature of an environment depends on the climate and water depth. Modern deep-sea environments maintain a steady temperature of a few degrees above 32°F (0°C), no matter what the latitude or climate at the surface. Temperature controls the solubility of calcium carbonate. Shallow seawater is supersaturated in calcium carbonate in warm climates but undersaturated in cool climates, so extensive deposits of limestone tend to be formed only in equatorial regions. Seawater also becomes undersaturated at depths of several thousands of meters in the oceans (below the lysocline), partly because of decreasing temperature but mainly because of an increased supply of carbon dioxide from the decay of organic detritus settling down from the surface layers of the oceans. As a consequence, sediments in the deepest parts of the ocean basins do not contain calcium carbonate, but consist of fine clay and siliceous planktonic skeletons (such as radiolarians). Environments sufficiently arid to evaporate large volumes of seawater are found only in hot climates, so temperature plays an important role in the formation of salt and gypsum deposits. At the other extreme, ice sheets and their associated sedimentary environments (moraines, glacial outwash fans, and so on) are found only in cold climates.

Salinity. The salinity of an environment depends on the aridity of the climate and on a source for the dissolved salts. Outside of restricted supersaline environments (playa lakes, coastal lagoons, and supratidal flats or sebkhas), most aqueous environments can be categorized as marine (with a salinity of 35 parts per thousand), fresh (with a salinity of less than

0.2 part per thousand), or brackish. Brackish waters are most frequently formed by the mixing of fresh water and seawater, for example, in estuaries, coastal lagoons, or deltas. Salinity in these environments frequently fluctuates widely, a condition very unfavorable to most forms of animals and plants. Therefore, brackish water environments tend to be characterized by their own specialized flora and fauna. Salinity is also important in determining the relative density of water masses. River water is less dense than seawater and tends to spread out over it as a surface layer when rivers enter estuaries or emerge from deltaic distributaries. A consequence is that river and seawater do not mix easily, and that a distinctive estuarine pattern of circulation is set up, with saline waters moving landward under surface fresh or brackish waters moving offshore. Within the ocean basins themselves much smaller differences in both salinity and temperature lead to stratification of water, with relatively little mixing between layers. The denser water masses generally move slowly along the bottom from the poles toward the equatorial regions and become progressively enriched in nutrients such as phosphates. Where the deeper waters are forced to the surface (zones of upwelling, which are generally found along the western equatorial margins of the continents), they may support prolific organic growth and produce sedimentary deposits rich in phosphates or organic matter. *See* UPWELLING.

Chemical factors. There are other chemical aspects of the environment that are of great importance for determining the mineralogy of the sediment. The two most commonly measured and cited indexes are the pH (hydrogen ion concentration, or relative acidity or alkalinity) and Eh (oxidation-reduction potential). Sea-water is mildly alkaline and oxidizing; river waters are generally mildly acidic and oxidizing. Highly acidic waters may be found in swamps and marshes, and highly alkaline waters in some shallow lakes in arid climates. Reducing waters can be found only in environments cut off from the atmosphere—for example, in the lower levels of some stratified water masses, such as in some lakes and estuaries. Reducing and acid conditions are, however, very commonly developed just below the sediment-water interface in muds deposited in both terrestrial and marine environments. Whatever the original color of such muds, they generally become black or gray below the surface as a result of the reduction of iron minerals to fine-grained iron sulfides. Conversely, sediment deposited on alluvial fans in arid regions remains in contact with air or oxygenated pore waters (in the vadose zone, above the water table) for long periods of time so that detrital iron-bearing minerals are oxidized to produce fine-grained red iron oxides. Thus the chemistry of the original depositional environment can be inferred from the sedimentary deposit only after the nature of the postdepositional (diagenetic) changes has been determined.

Biological factors. Besides the major environmental controls listed above, biological factors exert a very strong influence on the nature of many depositional environments. The compositions of the atmosphere and oceans are themselves strongly influenced by biological activity. There would be little or no oxygen in the atmosphere if it were not for the photosynthetic activity of plants. Deposition of calcium carbonate and silica in lakes and the oceans takes place largely through the action of plants and animals, and organic matter deposited along with mineral particles is largely responsible for the development of reducing conditions within sediments after deposition.

Vegetable material accumulates in swamps to form peat and coal, and fine organic detritus settles with marine muds and is the ultimate source of oil and gas. Both terrestrial and aquatic plants exert a trapping and binding action that tends to immobilize sedimentary particles, as, for example, when coastal sand dunes or tidal flats become stabilized by the growth of salt-tolerant grasses. Terrestrial vegetation plays an important role in rock weathering, and there is some evidence that less clay was produced by weathering in the period (pre-Silurian) before the land was colonized by plants. Mangrove swamps, coral reefs, and tropical rainforests are just a few examples of environments strongly influenced by a combination of a suitable climate and flora and fauna. The abundance of plants and animals in many environments is such that sediment deposited there may be strongly modified before ultimate burial. Sediment may be eaten for its organic content by deposit- or filter-feeding organisms, or may be disturbed by plant roots or burrowing animals and insects.

Ancient environments. In looking back over the geologic record of past environments, sedimentologists see some signs that ancient environments were not always the same as those on the modern Earth. There seem to have been periods of time (for example, Jurassic-Cretaceous) in which there were not large ice sheets, and thus no extensive glacial environments. Probably the bottom waters of the oceans were not as cold then as they are now, and there is some evidence that ocean circulation was less vigorous, so that stagnant, anoxic conditions could develop over large parts of some deep-sea basins.

Calcium carbonate is widely deposited in the ocean basins today (except in the deepest waters below the lysocline) largely because of the abundance of carbonate-secreting plankton, such as foraminifera. Such pelagic limestones are also common in some Cenozoic and Mesozoic sediments formed in deep water, but are rare in earlier deposits because such organisms did not evolve until Jurassic times. *See* FORAMINIFERIDA.

Conversely, some deposits abundant in the geologic record appear to be only sparsely represented by modern environments. Examples are the very extensive shallow-water limestones of the Paleozoic and marine evaporites. Some marine evaporites may have been formed by geologically unusual events, such as the complete desiccation of the Mediterranean Sea which has been suggested by some

geologists to explain the Miocene evaporites in that region.

Although the nature of carbonate-precipitating animals and plants, and therefore the nature of organic reefs and other environments of carbonate accumulation, has changed throughout geologic time, there remain some strong general resemblances between ancient and modern environments. The proportions of different environments have changed, but the main types of environments have remained the same. Only in Precambrian times is there evidence that sedimentary environments differed radically from those on the modern Earth. In the earliest Precambrian, volcanism seems to have been more active than at present and stable shallow marine environments less common. At the beginning of the Proterozoic, some 2 billion years ago, when shallow marine environments first became abundant, they were the site of extensive deposition of formations rich in iron. Modern analogs of these iron formations are hard to find, and many geologists believe that their formation is connected to the presence within the atmosphere of the time of only a very low abundance of oxygen. Such interpretations remain controversial. The guiding rule for most geologists attempting an environmental interpretation of ancient sedimentary rocks is actualism; that is, the present is the key to the past. *See* MARINE SEDIMENTS. Gerard V. Middleton

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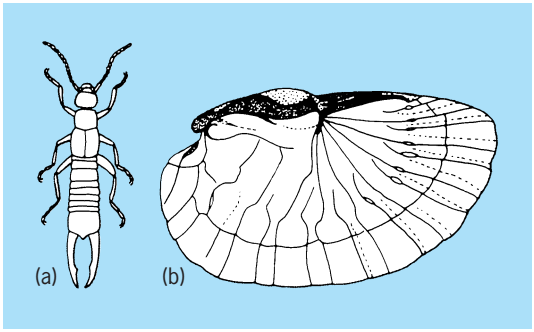
Dermaptera

An order of insects commonly known as earwigs, in reference to an unfounded superstition that they crawl into the ears of sleeping persons. Earwigs are elongate and flattened, with a pair of forcepslike appendages, or cerci, at the end of the abdomen (see **illus.**). They are 0.4–1.2 in. (10–30 mm) long, and have chewing mouthparts, gradual metamorphosis, long legs, and shortened, leathery forewings. Some species are flightless, while others are active fliers with well-developed hindwings that fold elaborately under the forewings when the insect is at rest. Most earwigs are brown or black.



Some features of Dermaptera. (*a*) Adult of Labidura. (*b*) Hindwing of *Forficula*. (*After C. P. Brues, A. L. Melander, and F. M. Carpenter, Classification of Insects, Harvard University Press, 1954*)

Earwigs are usually nocturnal and spend the daylight hours under bark or stones, in moist cracks and crevices, or in leaf litter at the soil surface. Some are attracted to lights. Most species scavenge on decaying animal or vegetable matter. A few species are predators on smaller insects. The widely introduced European earwig, *Forficula auricularia*, is an occasional pest of buildings, and crop plants, especially seedlings and floral crops.

Some earwig species display complex mating behavior in which the male grasps the female with his cerci. Females of some species show parental care, guarding eggs and nymphs against predators.

Worldwide, there are about 1100 species of earwigs, and most of these species are tropical. Of the 22 species known from North America, half are introduced. Some apparently arrived in sailing ships' ballast over 200 years ago. Fossil earwigs are known from the Jurassic Period. *See* INSECTA. David J. Horn

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Dermatitis

Inflammation of the skin with redness and scaling, or if acute, with blisters, edema, and formation of a crust; also known as eczema.

Human. Of the several distinct types of dermatitis, each has unique etiology, clinical characteristics, pathology, and treatment.

Atopic dermatitis. Atopic dermatitis usually begins in infancy but may continue into adult life. The eruption is characterized by red patches accompanied by intense itching. The patches become leathery and abraded with continued rubbing and scratching. The lesions often become secondarily infected, leading to moist discharge and crusting. Although any anatomic site may be affected, the classic locations in infants are the extremities, face, and scalp. In older children and adults, the inside of the elbows, the back of the knees, wrists, eyelids, and neck are most often involved.

Often familial, atopic dermatitis may be associated with asthma or allergic rhinitis. As with asthma and allergic rhinitis, about 80% of individuals with atopic dermatitis have elevated levels of immunoglobulin E (IgE). A wide range of other immunologic abnormalities have been found.

The disease has a genetic predisposition, but its expression is modified by environmental factors. The eruption may be made worse by low temperatures,

Some common causes of allergic contact dermatitis

Chemical or material	Product	Anatomic site
Nickel	Jewelry	Earlobes, neck, wrist
Paraphenylenediamine	Hair dye	Scalp, ears, face
Glyceryl monothioglycolate	Permanent-wave substances	Scalp, ears, face
Balsam of Peru	Fragrance	Eyelids, neck
Cinnamic alcohol	Fragrance	Eyelids, neck
Thimerosal	Contact lens solution	Eyelids
Quaternium-15	Preservative (cosmetics)	Face, hands
Imidazolidinyl urea	Preservative (cosmetics)	Face, hands
Thiuram, mercapto	Rubber	Hands, feet
Potassium dichromate	Cement, leather, metals	Hands, feet, generalized
Formaldehyde	Nail polish, clothing	Eyelids, generalized
Acrylates	Artificial nails	Eyelids
Cinnamic aldehyde	Toothpaste	Perioral
Colophony	Adhesives	Feet, hands
Neomycin, cortisones, benzocaine	Topical medications	Wherever applied

decreased humidity, contact with wool, immunizations, and upper respiratory infections. Eggs, peanuts, milk, wheat, fish, soy, and chicken may accelerate the disease. Nickel, either in direct contact with the skin or via dietary intake, may play a role in the disease.

Individuals with atopic dermatitis have an increased susceptibility to certain infections. Over 90% of eczematous lesions contain *Staphylococcus aureus*, and many individuals are colonized with the organism. Widespread fungal infections and warts occur. Secondary infection with *Herpes simplex* virus, known as eczema herpeticum, is perhaps the most serious complication of atopic dermatitis. *See* HERPES; STAPHYLOCOCCUS.

A skin biopsy in atopic dermatitis shows epidermal and dermal changes. In earlier lesions, there is mild thickening of the epidermis with intercellular edema, and superficial dermal inflammation. With chronic rubbing, the edema becomes less prominent and thickening of the epidermis is more marked.

Treatment of atopic dermatitis includes oral antihistamines, emollients, topical corticosteroids (particularly ointments), and for more severe cases, systemic corticosteroids. Avoidance of irritants (wool, chemicals, harsh soaps and detergents, perfumes), extremes in temperature and humidity, overbathing, and certain foods is imperative. Mild cleansers are preferred. For weeping and crusted lesions, warm compresses with tap water or aluminum acetate solution (Burow's solution) are helpful. When secondary infection by *S. aureus* exists, oral antibiotics or topical mupirocin, an antibacterial ointment, should be used. *See* ANTIHISTAMINE; STEROID.

Contact dermatitis. Contact dermatitis is a reaction that causes acute itching, redness, swelling, large blisters, and in chronic cases, red, scaly papules (raised bumps) and plaques (abnormal flat areas). The two forms are irritant contact dermatitis and allergic contact dermatitis. In irritant contact dermatitis the eruption is caused by a nonallergic reaction resulting from exposure to an irritating substance. Allergic contact dermatitis is an immunologic reaction in persons who have been previously exposed (sensitized) to the allergen, the classic example being poison ivy. Linear streaks and geometric shapes suggest contact dermatitis. The lesions occur 24–48 h after exposure to an irritant or allergen.

The anatomic distribution of the skin lesions is important in determining the etiologic agent. Common irritants include solvents, cutting oils, petroleum products, plastics and resins, food products, and soaps and cleaners (see **table**). The diagnosis of contact dermatitis is made by a careful history, physical examination, and in some cases, patch testing to identify the allergen.

In acute contact dermatitis the histopathology shows edema with formation of blisters and sparse dermal inflammation. Over time, there is thickening of the epidermis and greater inflammation. Increased exposure to irritants may also cause localized death of skin tissue and ulceration.

The key to treatment is removal of the offending agent. Antihistamines, emollients, and topical and systemic corticosteroids are helpful. For treatment of cases of acute allergic contact dermatitis (for example, poison ivy), most individuals require oral prednisone.

Stasis dermatitis. Scaly, reddened, ill-defined plaques on the lower legs, often accompanied by swelling, hyperpigmentation, loss of hair, or ulceration, are common symptoms of stasis dermatitis. The changes are associated with continued poor venous circulation. Thickening of the outermost layer of the epidermis (hyperkeratosis), edema, and thickening of the epidermis occur. Within the outermost layer of the dermis, there is inflammation, vascular changes, hemorrhage, and iron pigment. Treatment consists of improving circulation, reducing edema (by means of elevation, compression, and diuretics), and administering topical corticosteroids.

Seborrheic dermatitis. Seborrheic dermatitis is characterized by reddish to yellow greasy scale that frequently forms on the face (eyebrows, nasolabial folds, forehead), ears, scalp, upper back, central chest, and anogential region. In contrast to atopic and contact dermatitis, itching is uncommon.

The etiology of seborrheic dermatitis is multifactorial. Genetic and hormonal factors, emotional stress, and the fungus *Pityrosporum ovale* have all been

proposed. Skin biopsies of individuals with seborrheic dermatitis reveal thickening of the epidermis, mild edema, the presence of neutrophils within the outermost layer of the epidermis, and superficial inflammation of the dermis. *See* FUNGI.

Topical corticosteroids are the mainstay of treatment for seborrheic dermatitis. In cases in which *P. ovale* is involved, topical antifungal agents (for example, ketoconazole cream) are useful. Selenium sulfide, tar, zinc pyrithionate, ketoconazole, and resorcin shampoos are useful therapies for scalp involvement. Topical or oral antibiotics are required in cases with secondary bacterial infection. In generalized or recurring cases, systemic corticosteroids may be necessary.

Exfoliative erythroderma. Exfoliative erythroderma is characterized by widespread, warm redness and scaling. Nail degeneration and loss, hair loss, fever, chills, and enlargement of the lymph nodes may also occur. The eruption may be due to an underlying primary skin disorder, such as dermatitis (seborrheic, contact, or atopic) or psoriasis; drug allergy; or leukemia or lymphoma, particularly cutaneous T-cell lymphoma. A careful history, physical examination, and skin biopsy are helpful in determining the etiology. *See* LEUKEMIA; LYMPHATIC SYSTEM; LYMPHOMA.

Individuals with exfoliative erythroderma are best managed in a hospital setting. Because heat, water, and protein losses occur through damaged skin, careful monitoring of fluid balance, protein losses, and body temperature is imperative. The application of wet dressings or topical corticosteroids and administration of oral antihistamines are important aspects of treatment. Systemic corticosteroids are useful if dermatitis or a drug-induced eruption is involved, but they may exacerbate psoriasis. Systemic antibiotics should be used in cases where a secondary bacterial infection has developed. All other medication should be discontinued if possible. *See* INFECTION.

Lisa M. Cohen; Kenneth A. Arndt

Flea allergy dermatitis. Flea allergy dermatitis, sometimes called flea-bite dermatitis, affects the skin of small pets (dogs, cats, and ferrets), and results from an allergic reaction to protein substances deposited on (or under) the surface of the skin at the time of flea feeding. It is by far the most common cause of allergic dermatitis in pet animals and a major cause of itchy skin in geographic regions where fleas are found.

Pathogenesis. The most common species of flea that parasitizes dogs and cats is *Ctenocephalides felis felis.* An allergic reaction may occur in dogs of any age following the exposure to the flea saliva. This sensitization does not require a genetic predilection as do other types of allergies, but may be antibody induced (immunoglobulin E), cell mediated (delayed reaction), or brought about by other mechanisms. Dogs only several weeks old can develop an allergic reaction to fleas, while other dogs or cats can live with fleas for many years before developing the allergy. *See* ANTIBODY.

The flea remains on the host, leaving only to find a new, more desirable host; it rarely returns to the environment. Since the adult flea is geared primarily for perpetuation of the species, the female consumes quantities of blood to sustain nourishment for egg production, producing 1000–1500 eggs over a period of 100 days.

Clinical signs. Continuous or intermittent exposure to fleas perpetuates the clinical signs, the most prominent being itchy skin. Even a limited flea infestation may elicit profound clinical signs. On dogs, red, scaly skin over the top of the lower back by the tailhead appears with a typical pattern of self-inflicted hair loss as well as lesions on the abdomen and rear legs. Secondary *Staphylococcus* infections are common in chronic cases that do not receive appropriate treatment. Scales and crusts are often observed associated with a variable rash. Cats may have a more varied array of clinical symptoms. Focal crusts (miliary dermatitis) develop on the lower back, tailhead region, and thighs. Comparable lesions may be found around the neck. Often, hair loss with minimal dermatitis affecting the abdomen is caused by excessive licking and hair pulling in that area. The disease may progress to a generalized pattern, but usually spares the head, feet, and forelegs. Self-inflicted trauma to the skin is produced by removal of the fleas through scratching, biting, licking, and chewing.

Diagnosis. The diagnosis of a flea allergy is based on the history of the animal and on clinical examination. Confirmation is made by obtaining a positive reaction to a commercial flea antigen injected in the skin, although this may be observed in only 90% of the cases. Response to therapy effective for flea control provides a tentative diagnosis. Finding evidence of fleas may not be possible. Identification and removal of fleas is best accomplished by using a fine-toothed comb, with 24–26 teeth per inch, for a short- or medium-coated pet. Small specks of black material present on the skin surface may represent flea dirt (feces), which will turn red (representing the whole blood ingested by the flea) if applied to a water-moistened cottonball and allowed to stand for 5–8 min.

Treatment. Treatment of flea allergy is directed at three areas: flea control, which includes removal of the flea from the infested pet, treatment of associated animals, and aggressive treatment of the animals' habitat; symptomatic treatment for the itching; and treatment for secondary complicating factors such as bacterial skin infection (pyoderma) or dry skin.

A variety of products are commercially prepared for animal treatment: topical products include shampoos, pour-on insecticides, sprays, powders, and spot treatments and collars; systemic treatments include oral medications and spot treatments, where percutaneous absorption provides systemic concentration of the insecticide, which the flea ingests during feeding. Most topical treatments have a short duration of activity with poor residual killing or repellent effectiveness. Systemic products also have limited activity for the flea-allergic animal because

the allergic reaction is induced by flea feeding. Some products (organophosphates and carbamates) may be toxic with repeated applications. Safer ingredients include the botanicals derived from plant sources. Examples include pyrethrins (chrysanthemum), rotenone (derris plant), or *d*-limonene or linalool from citrus produce (orange peel).

Insect growth regulators are successful because they interrupt the flea life cycle in preadult forms (eggs and larvae). Although they are principally used for premises treatment, some products have been developed for application on the animal or systematic administration.

Adjunctive treatment is directed toward topical therapy to reduce the dryness, systemic and topical drugs to decrease the itching, and specific drugs for treating concurrent diseases such as bacterial skin infections. Systemic anti-inflammatory steroid treatment may be necessary if the itching is intense and aggressive biting, licking, chewing, and scratching are present. Long-term therapy with steroids requires proper selection and interval of therapy to avoid the many potential side effects of these drugs. Antihistamines may provide some relief for itchy skin, but have a less predictable effect than steroids. However, neither antihistamines nor steroids will be effective if flea infestation on the flea-allergic pet continues. *See* ALLERGY; SIPHONAPTERA; SKIN DISORDERS.

J. M. MacDonald

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Dermoptera

An order of mammals comprising one family (Cynocephalidae), one genus (*Cynocephalus*), and two species—*C. variegatus* (Malayan colugo) and *C. volans* (Philippine colugo). These mammals inhabit the rainforests of Southeast Asia from the Malay Peninsula and Thailand to the Philippine Islands, Java, and Borneo.

General characteristics. Colugos are about the size of a large squirrel or domestic cat. Their most distinctive feature is a gliding membrane (patagium) that attaches to the neck and the sides of the body. It extends from the tips of the fingers and toes to the tip of the tail. It gives the appearance of an animal wearing a fur coat that reaches to its toes. When at rest, the furry membrane hangs in loose, voluminous folds. No other volant mammal possesses such an extensive flight membrane. Although colugos are also known as flying lemurs, they cannot fly. (Bats, order Chiroptera, are the only flying mammals.) By spreading the patagium, the colugo is able to perform controlled glides of 100 m (330 ft) or more with little loss of height. The head is broad with a blunt muzzle and short, rounded ears. The arms, legs, and tail are long and slender. The limbs terminate in strong, sharp, recurved claws. These nocturnal mammals have large, prominent eyes and stereoscopic vision. The single pair of mammae in females are located at the sides of the body, almost in the armpits. The flight membrane and the dorsal surface are mottled brownish-gray with males being more brownish. Scattered white spots are present on the dorsal surfaces of the Malayan colugo. The mottled coat blends with the bark of trees. The underparts are paler. The dental formula is I 2/3, C1/1, PM 2/2, M 3/3 × 2 for a total of 34 teeth. The lower incisors have been modified into "comb teeth" with as many as 20 prongs radiating from one root. These specialized teeth are unlike those in any other mammal and may serve as food strainers, scrapers, and/or as an aid in grooming. Adults have head-body length of 340–420 mm (13–16.5 in.), a tail of 175–270 mm (7–11 in.), and a weight of 1.0–1.75 kg (2–4 lb). The span of the patagium is 70–120 cm (2.3–4.0 ft).

Malayan colugo (*Cynocephalus variegatus*), also known as the Malayan flying lemur. (*Photo by Ampika Korcharernkit; courtesy of Suwit Punnadee, Director of the Gibbon Rehabilitation Project, Wild Animal Rescue Foundation of Thailand*)

Habitat. Colugos are found in both lowland and mountainous areas in Southeast Asia, where they inhabit primary and secondary forests as well as coconut groves and rubber plantations. They are totally arboreal and seldom, if ever, descend to the ground. They are unable to stand erect when on the ground and are clumsy climbers. When feeding, roosting, or moving about in the trees, they do so in an upside-down position like a sloth.

Behavior. These nocturnal mammals spend the daylight hours either in a hole or cavity of a tree or hidden in dense foliage. The diet of these herbivores consists primarily of foliage such as leaves, shoots, buds, and flowers. Since leaves have a relatively low nutrient content, large quantities must be consumed.

Soft fruits may occasionally be eaten. Sufficient water is obtained through food and by licking wet leaves. Adults are normally solitary but have broadly overlapping home ranges. They are highly territorial, not only with respect to their sleeping place, but also with respect to their foraging trees.

Reproduction. Breeding may occur throughout the year with gestation being approximately 60 days. Most litters consist of a single young. The newborn is very undeveloped, similar to a marsupial, and is carried on the mother's belly, even when she is gliding. Near the tail, the mother's patagium can be folded to form a pocket for the young. Weaning occurs at about 6 months of age. It may require 3 years for colugos to reach adult size.

The slow rate of reproduction and maturation along with loss of forest habitat through logging and conversion to agricultural use has led to the Philippine colugo being classified as vulnerable by the International Union for the Conservation of Nature (IUCN). This species is also hunted for its meat and fur. The major natural predator is the Philippine monkey-eating eagle (*Pithecophaga jefferyi*).

Donald W. Linzey

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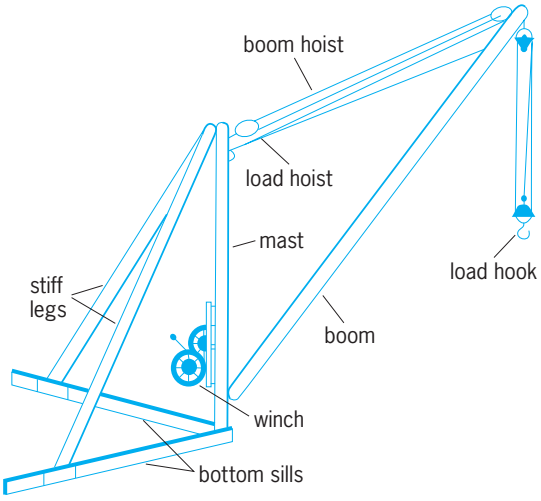
Derrick

A hoisting machine consisting usually of a vertical mast, a slanted boom, and associated tackle (see **illus**.). Derricks have a wide variety of forms. The mast may be no more than a base for the boom; it may be a tripod, an A-frame, a fixed column, and so on. Fixed stays may guy it in place. The boom may be fixed, it may pivot at the base of the mast, it may swing horizontally from near the top of the mast, or it may be omitted. The derrick may be permanently fixed, temporarily erected, or mobile on a cart or truck.



A boom derrick equipped with a swinging mast and anchored legs.

Derricks are widely used in construction, in cargo handling, and in shops. Their lifting action is intermittent compared to bucket conveyors, and their coverage is limited by the reach of the boom. A manual or powered winch provides the lifting action by coiling in the running tackle, the load swinging free as it rises. *See* BULK-HANDLING MACHINES; HOISTING MACHINES; OIL AND GAS WELL DRILLING.

David O. Haynes; Frank H. Rockett

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Derris

A genus of tropical shrubs belonging to the legume family (Leguminosae). These plants, with their long branches climbing over other vegetation, occur as members of the jungle undergrowth in Malaysia. Extracts of the roots of *Derris elliptica* have long been used by the natives as an arrow poison and to stupefy fish so they can be caught more easily. Derris root is an excellent insecticide, being harmful to both chewing and sucking insects, but not poisonous to human beings. The insecticidal ingredient of derris root is a white, crystalline substance, which is called rotenone. *See* ROSALES.

Perry D. Strausbaugh; Earl L. Core

Descriptive geometry

A mathematical-graphical procedure that has for its purpose the visualization of structures and their exact representation in drawings. After analysis of the

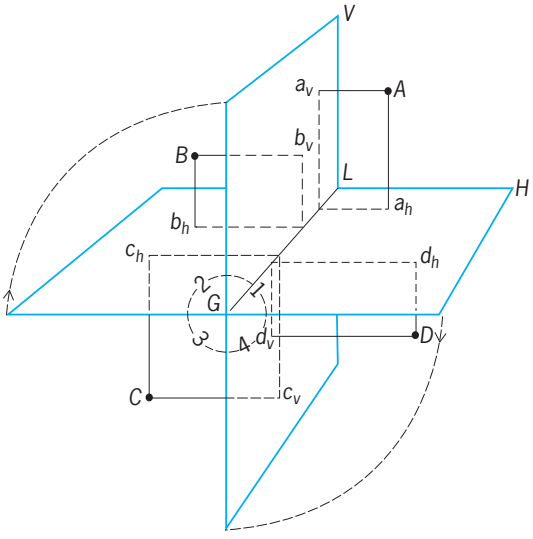

Fig. 1. Planes of projection. (*After G. J. Hood and A. S. Palmerlee, Geometry of Engineering Drawing, 4th ed., McGraw-Hill, 1958*)

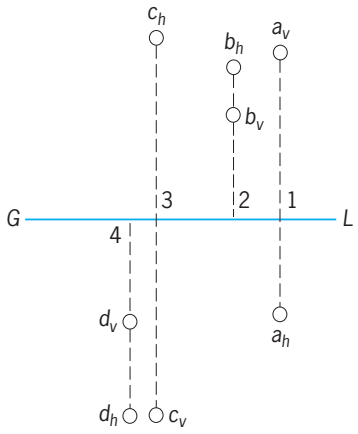

Fig. 2. Projection of points onto projection planes. (*After G. J. Hood and A. S. Palmerlee, Geometry of Engineering Drawing, 4th ed., McGraw-Hill, 1958*)

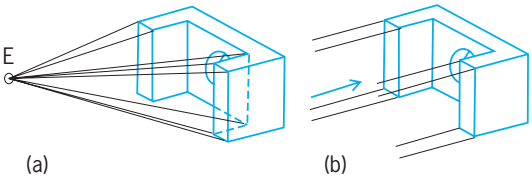

Fig. 3. Methods of viewing objects. (*a*) Perspective viewpoint (converging rays). (*b*) Orthographic viewpoint (parallel rays). (*After G. J. Hood and A. S. Palmerlee, Geometry of Engineering Drawing, 4th ed., McGraw-Hill, 1958*)

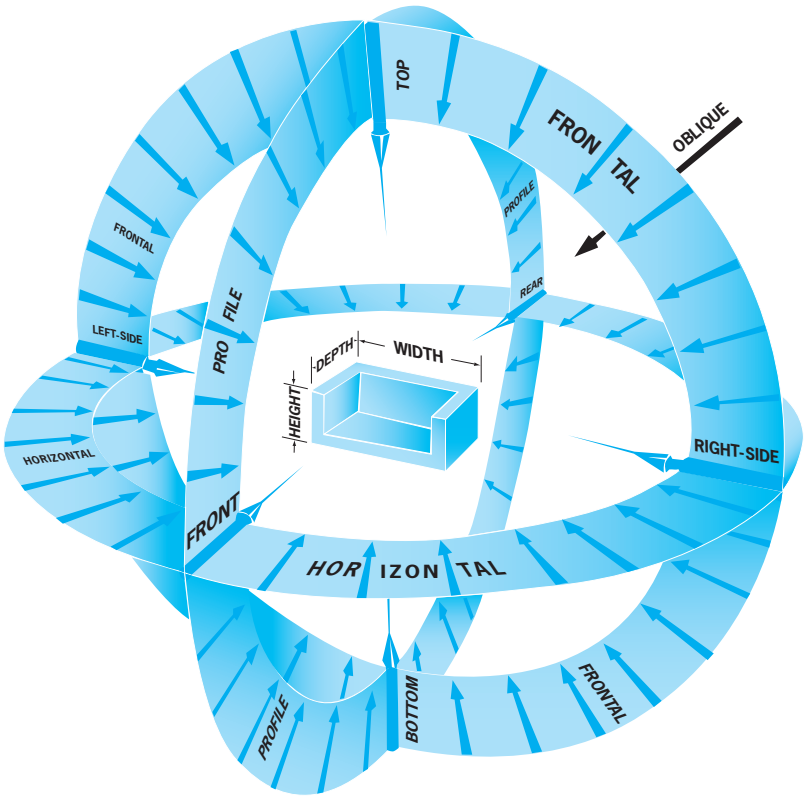

Fig. 4. A series of viewing positions for auxiliary views. (*After G. J. Hood and A. S. Palmerlee, Geometry of Engineering Drawing, 4th ed., McGraw-Hill, 1958*)

structure, each element is shown in the drawing in its exact geometrical relation to the other elements.

There are two basic methods of descriptive geometry: the projection method and the direct method. The two methods differ as regards the atttitude of mind toward the structure and toward the drawing that represents the structure.

Projection method. Gaspard Monge (1756–1818), a French mathematician, originated the projection method of descriptive geometry. While working as a designer for the French government, he was given the job of making plans for a proposed fortress. This was a tedious process and involved long calculations. He invented graphical solutions and completed the plans in such a short time that at first the commandant refused to accept them. For a long time the graphical process was kept a state secret, finally being revealed about 1795.

In the projection method the horizontal projection plane H and the vertical projection plane V intersect in the line GL, which is called the ground line (**Fig. 1**). These two projection planes divide space into four quadrants, or angles, as numbered in Fig. 1.

Point A, in the first quadrant, is projected onto the horizontal plane at a_h by means of a projection line perpendicular to the H plane, and onto the vertical plane at a_v by means of a projection line perpendicular to the V plane. The projections of the points B, C, and D in the other quadrants are located in a similar manner. If desired, a profile projection plane, perpendicular to the H and V planes, may be introduced. Right-angle projection, as described above, is called orthographic projection.

To represent horizontal and vertical projections on a flat sheet of paper, the planes are conceived as being hinged along the ground line and brought together by closing the second and fourth quadrants. Projections of A, B, C, and D then appear in a single plane (**Fig. 2**). The H and V projections of a point are always in the same perpendicular to the ground line. The usual custom in the United States is to draw objects as if they had been projected from a position in the third quadrant; this is called third-angle projection. First-angle projection is standard in some countries and professions.

There are two general types of views, perspective and orthographic (**Fig. 3**). A perspective view of an object is observed from a fixed station point, or point of view, by means of converging rays of light that meet at the eye of the observer. An orthographic view of an object is observed in a chosen direction by means of parallel rays of light. As shown in Fig. 3*b*, the observer may be imagined at a distance considered to be infinity.

Direct method. In the direct method the attention is focused on the visualized structure or object. Each view of the object is obtained by looking at the object in a definite direction. The view is orthographic. A view never is considered as two-dimensional or as projected or drawn on a plane. This is the way the engineer who makes and reads the drawings thinks of views.

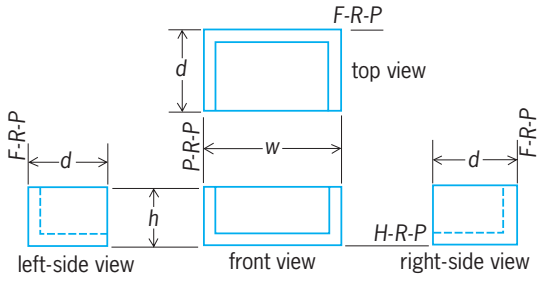

Fig. 5. Principal views of an object and some reference planes. (*After G. J. Hood and A. S. Palmerlee, Geometry of Engineering Drawing, 4th ed., McGraw-Hill, 1958*)

Orthographic views may be classified into three types: principal views, auxiliary views, and oblique views. The object can be viewed from any direction around three rings—horizontal, frontal, and profile (**Fig. 4**). The rings represent three mutually perpendicular planes. The intersections of the rings define three mutually perpendicular directions from which six principal views are observed: front and rear, top and bottom, right and left sides.

An auxiliary view can be observed around any ring in a direction perpendicular to one, and only one, of the directions in which principal views are observed.

All views other than principal or auxiliary views are oblique. In Fig. 4 a single arrow, marked oblique, indicates one of the infinite number of directions in which oblique views are observed.

Dimensions of structures. All structures occupy space and have three dimensions: width, height, and depth. These are measured in three mutually perpendicular directions (Fig. 4).

Grouping of views. In **Fig. 5** are shown four views of the object pictured in Fig. 4. Each view is placed in relation to its adjacent view in the position from which it was viewed. Edge views of three reference planes are indicated in Fig. 5. For example, *H-R-P* is the edge view of a horizontal reference plane from which all vertical height dimensions of the object may be measured. Similarly, all depth dimensions are measured from the frontal reference plane *F-R-P*, the edge view of which is seen in the two side views and in the top view.

Reference planes. **Figure 6** shows the front and top views and four auxiliary elevations of a bearing. In the front view a horizontal reference plane is taken in the auxiliary elevations. The height dimension, or elevation, of each point in the front view, above or below the reference plane, is measured in the front view and then is transferred to each of the auxiliary elevations.

In practice, views are drawn from orientations that show the true sizes of selected lines or surfaces; that is, views are drawn perpendicular to such lines or surfaces. Thus, a triangular pyramid might be viewed from four directions, each of which shows one face in its true shape.

Reading the drawing. The object in a drawing is regarded as stationary. As the reader looks at first one view and then another, he gradually obtains a detailed mental picture of the object. The reader imagines that he moves around the object, viewing it successively in each direction indicated by the view.

Geometrical elements of structures. Points, lines, and surfaces are the geometrical elements of

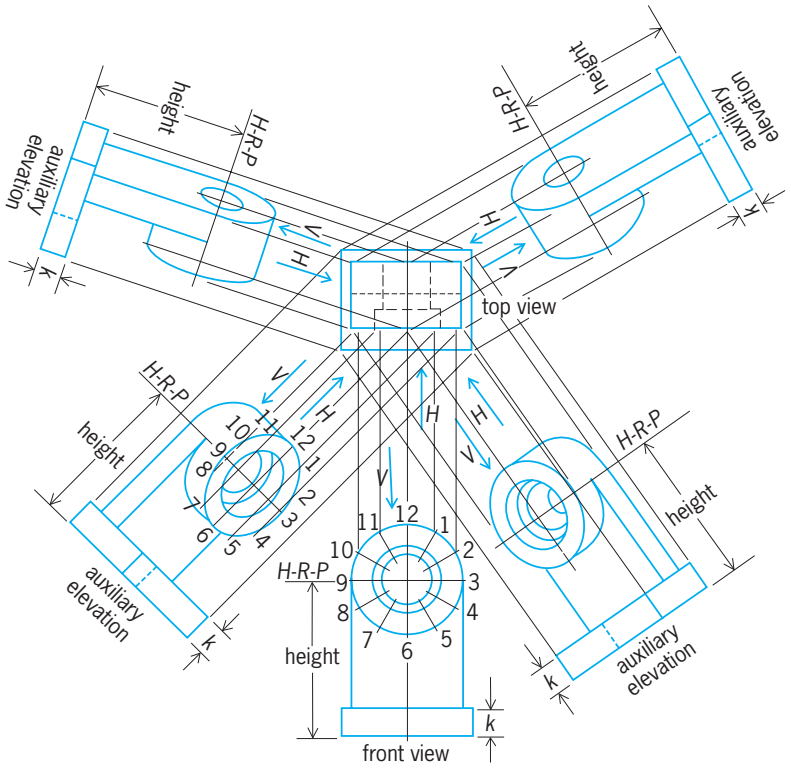

Fig. 6. Auxiliary views of a bearing. (*After G. J. Hood and A. S. Palmerlee, Geometry of Engineering Drawing, 4th ed., McGraw-Hill, 1958*)

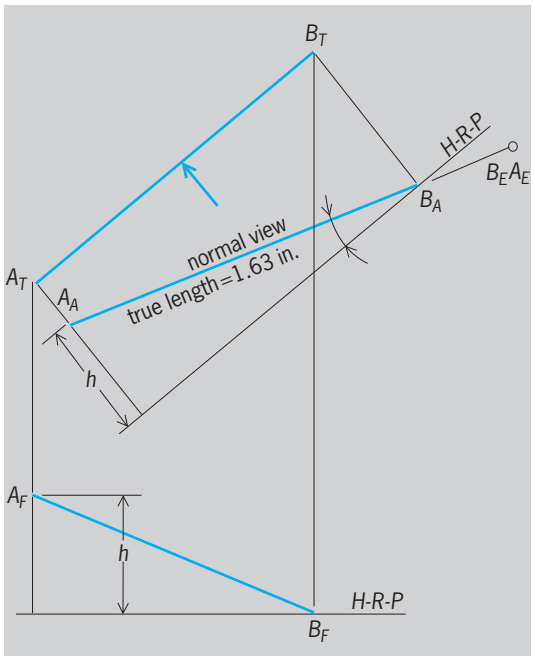

Fig. 7. True length of a line, seen in a normal view. (*After G. J. Hood and A. S. Palmerlee, Geometry of Engineering Drawing, 4th ed., McGraw-Hill, 1958*)

Geometrical elements of structures

- *Points*
- *Lines*
  - *Straight*
  - *Curved*
    - *Single-curved*: Circle, Ellipse, Parabola, Hyperbola, Trochoid, Spiral, Involute, Cycloid, Epicycloid, Hypocycloid, Sinusoid
    - *Double-curved*: General type, Helix
- *Surfaces*
  - *Plane*: Triangle, Quadrilateral, Polygon, Prism, Wedge, Pyramid, Regular polyhedron
  - *Ruled*
    - *Single-curved*: Cylinder, Cone, Convolute
    - *Warped*: General type, Hyperbolic paraboloid, Conoid, Helicoid, Hyperboloid of one sheet, Cylindroid
  - *Double-curved*: General type, Sphere, Ellipsoid, Paraboloid, Hyperboloid of two sheets, Torus, Surfaces of revolution, Serpentine

structures. Points are elements of lines, and lines are elements of surfaces. Lines are generated by a moving point; surfaces are generated by a moving line. The law that governs the motion of the moving point or line determines the nature of the generated line or surface. In the **table** of geometrical elements of structures, various kinds of lines and surfaces are named.

The first step in preparing a descriptive diagram is to decide what views are necessary to show the geometrical relations and locations of all elements and parts of the object. To draw the views, it is necessary to understand the measurements that must be made in each view; the principles involved in showing parallel, perpendicular, and angular relations between elements; the methods of showing the true lengths of lines and the true shapes of plane surfaces; the intersections of surfaces; and the development of surfaces.

True length of line. A normal view of a line or plane is taken in a direction perpendicular to the line or plane (**Fig. 7**). The true length of the line *AB* is seen in the normal view, here taken in a horizontal direction that is perpendicular to the line *B* in the top view. Dimension *b* is obtained from the front view. The inclination of line *AB* to the horizontal also is indicated.

Parallel lines. Parallel lines appear parallel in every view, as is illustrated by a parallelogram (**Fig. 8**).

Perpendicular lines. Perpendicular lines appear perpendicular only in a view that is normal to one or both of the lines (**Fig. 9**). In Fig. 9, *RC* is a given line, but only the top view of the line *RB* is given. The front view of *RB* is to be found. The numerals indicate the consecutive order in which the lines are drawn. A normal auxiliary elevation is drawn

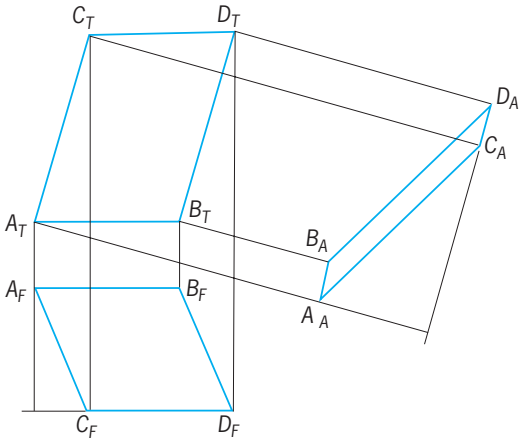

Fig. 8. Parallel lines in front, top, and auxiliary views. (*After G. J. Hood and A. S. Palmerlee, Geometry of Engineering Drawing, 4th ed., McGraw-Hill, 1958*)

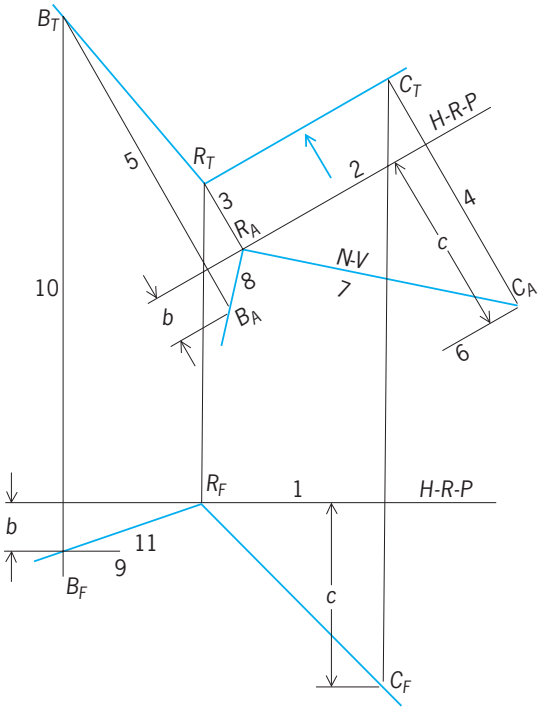

Fig. 9. Perpendicular oblique lines. (*After G. J. Hood and A. S. Palmerlee, Geometry of Engineering Drawing, 4th ed., McGraw-Hill, 1958*)

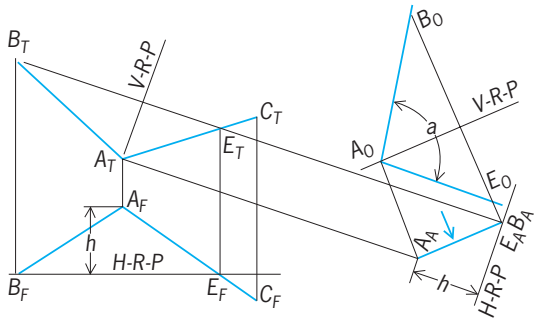

Fig. 10. Angle between two skew lines. (*After G. J. Hood and A. S. Palmerlee, Geometry of Engineering Drawing, 4th ed., McGraw-Hill, 1958*)

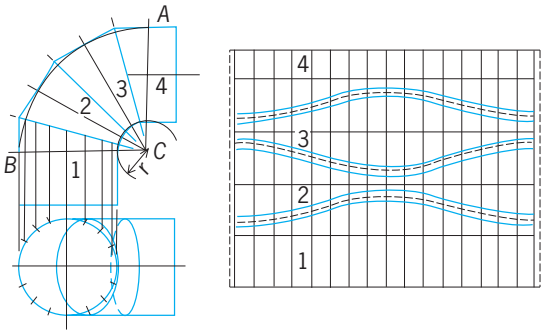

Fig. 11. Layout and top of front views of a four-piece elbow. (*After G. J. Hood and A. S. Palmerlee, Geometry of Engineering Drawing, 4th ed., McGraw-Hill, 1958*)

perpendicular to the line *RC* in the top view. In this normal view, line *RC* is drawn and *RB* is drawn perpendicular to *RC*. This view now shows a normal view of *RC*, but not of *RB*. The front view of *RB* may now be drawn.

Angle between lines. The angle between lines *AB* and *AC* is to be found (**Fig. 10**). An auxiliary elevation taken in the direction of horizontal line *BE* is drawn. This is an edge view of the plane containing given lines *AB* and *AC*. Next, a view taken in a direction perpendicular to this plane shows a normal view of the two given lines and the true size of the angle between them.

Intersection and development of surfaces. A structure is limited by the surfaces that bound it. Each surface of the structure is in turn limited by the straight or curved lines in which it intersects adjacent surfaces. These lines of intersection determine the edges and joints of the structure and are shown in drawings to describe the structure. In addition to the lines of intersection, the outlines, or contours, of curved surfaces also are shown in drawings. All intersections must be accurately located to make accurate developments.

For example, sheet metal bends or folds along a straight line; hence only plane-faced and single-curved surfaces may be developed. The single-curved surfaces are cylinders, cones, and convolutes. Warped surfaces and double-curved surfaces cannot be developed but must be formed by stretching the metal. Each type is useful in engineering designs.

A four-piece elbow is constructed from parts of right circular cylinders of uniform cross section (**Fig. 11**). At the right the development of each section is outlined by the solid lines. Allowance for seams is indicated by broken lines. No material is wasted.

The warped surfaces, such as the hyperbolic paraboloid, the concoid, the helicoid, hyperboloid of revolution, and others, have definite uses and are relatively easy to construct if the geometry of the surfaces is understood.

Topographical problems. Engineers encounter topographical problems dealing with land contours, underground and surface workings and surveys, highways, bridge piers, dams, foundations, retaining walls, excavations, underground and surface water supplies, drainage, irrigation, geological formations, rock strata, oil wells, mines, tunnels, veins of ore, dumps, layout of grounds, landscaping, and structures built into the ground and on the surface of the ground (**Fig. 12**).

The upper half of Fig. 12 illustrates the map of a section of mountainous country. The section has been prospected, and outcrops or traces of veins of ore have been located. Items such as the strike, dip, thickness, line of outcrop, and overburden of the veins can be determined and the location and length

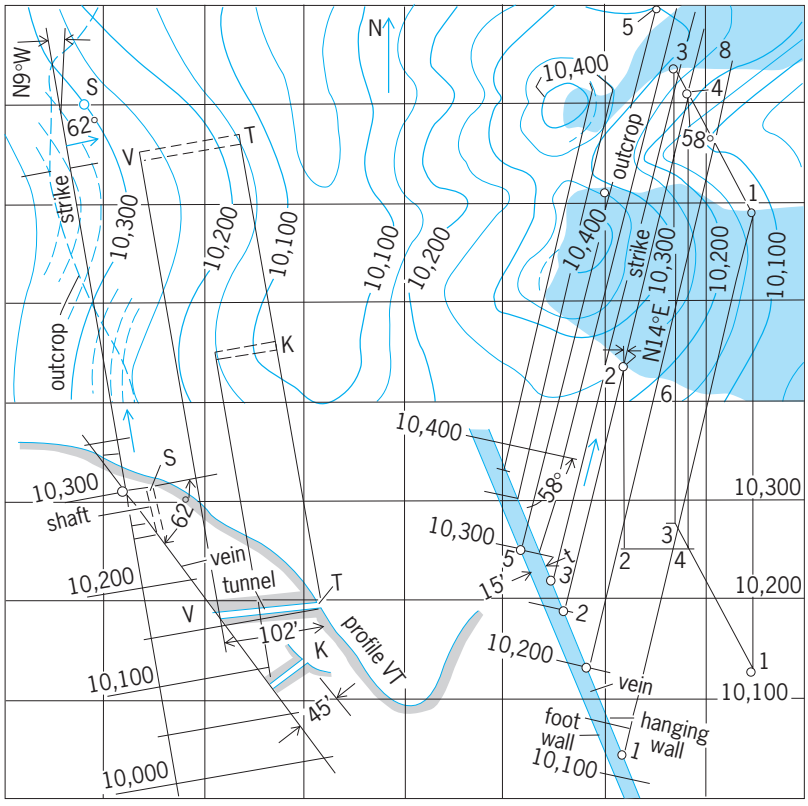

Fig. 12. Veins and outcrops of a mountainous section shown in top half. The bottom half provides additional data. (*After G. J. Hood and A. S. Palmerlee, Geometry of Engineering Drawing, 4th ed., McGraw-Hill, 1958*)

of tunnels and shafts established. *See* DRAFTING; ENGINEERING DRAWING.

Albert S. Palmerlee; Charles J. Baer

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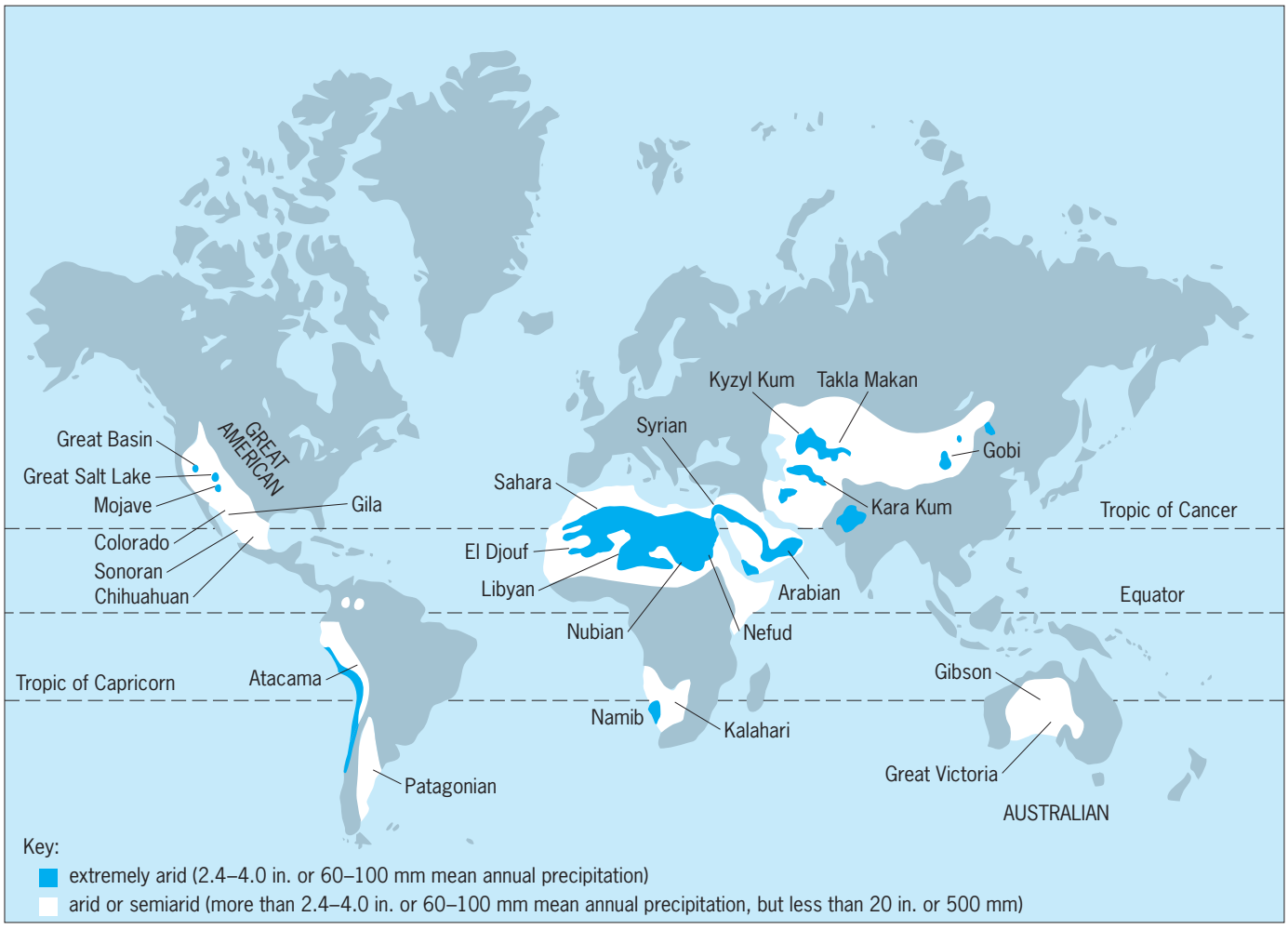
Desert

No precise definition of a desert exists. From an ecological viewpoint the scarcity of rainfall is all important, as it directly affects plant productivity which in turn affects the abundance, diversity, and activity of animals. It has become customary to describe deserts as extremely arid where the mean precipitation is less than 2.5–4 in. (60–100 mm), arid where it is 2.5–4 to 6–10 in. (60–100 to 150–250 mm), and semiarid where it is 6–10 to 10–20 in. (150–250 to 250–500 mm). However, mean figures tend to distort the true state of affairs because precipitation in deserts is unreliable and variable. In some areas, such as the Atacama in Chile and the Arabian Desert, there may be no rainfall for several years. It is the biological effectiveness of rainfall that matters and this may vary with wind and temperature, which affect evaporation rates. The vegetation cover also alters the evaporation rate and increases the effectiveness of rainfall. Rainfall, then, is the chief limiting factor to biological processes but intense solar radiation, high temperatures, and a paucity of nutrients (especially of nitrogen) may also limit plant productivity, and hence animal abundance. *See* PRECIPITATION (METEOROLOGY).

The main desert regions of the world are shown in the **illustration**. Most lie within the tropics and hence are hot as well as arid. The Namib and Atacama coastal deserts are kept cool by the Benguela and Humboldt ocean currents, and many desert areas of central Asia are cool because of high latitude and altitude.

Plant production. As a consequence of unreliable rainfall, plant productivity is much more variable than in most ecosystems. It may vary from none to 880 dry lb/acre · yr (1000 dry kg/ha · yr) although in most places there is usually some



The deserts of the world; those within the tropics are hot as well as arid.

productivity even when there is no rainfall. The average biomass is generally low at 0.004–0.144 dry lb/ft^2 (0.02–0.7 dry kg/m^2) compared to 9.25 dry lb/ft^2 (45 dry kg/m^2) in tropical forest and 6.17 dry lb/ft^2 (30 dry kg/m^2) in temperate forest. Another feature of desert vegetation is the low percentage of green, photosynthetic plant biomass. There is also about three times as much dead as living plant material, a high figure compared to most other ecosystems. *See* BIOLOGICAL PRODUCTIVITY; BIOMASS.

Between 10 and 20% of the living plant material is eaten by consumers, a figure typical of terrestrial ecosystems, where the level of consumption may be as low as 5% (compared to over 90% in ocean ecosystems). Much desert vegetation tends to be hard and prickly and hence unpalatable to many consumers. Dead plant material does not accumulate and is utilized by the decomposers, particularly wood-eating termites. Seeds are abundant and can survive a long time without germinating. Many animals, including rodents, ants, and birds, depend heavily upon them for food. There is evidence of both intra- and interspecific competition for seeds, the production of which varies markedly with variations in rainfall.

A common viewpoint is that desert plants and animals live in a harsh environment and as a consequence have evolved many morphological and behavioral adaptations that enable them to "escape" the rigors of their surroundings. Certainly extreme aridity coupled with extreme heat seems a stressful environment, but whether desert organisms really possess more escape adaptations than, say, rainforest organisms is a matter for conjecture. Nevertheless, specific adaptations to drought and heat can be readily identified.

Adaptations in plants. In many deserts, annual rather than perennial plants form the bulk of the climax vegetation, quite unlike the situation in most terrestrial ecosystems where perennial plants dominate. Deserts, especially those within the tropics, periodically "bloom" with flowering annuals soon after a significant rainfall. In general, the more unreliable the rainfall in a desert, the greater the abundance and diversity of annual plants. Estimates vary, but these annuals account for up to half the primary production in a year, although for a given site they may fail altogether if there is no rain. Many annuals last for only a very short time. They tend to have the C_4 photosynthetic pathway, which means they grow quickly and opportunistically. It is these plants in particular that produce the vast quantities of seed which under extreme aridity can last for years, providing a "seed bank" until conditions are right for germination. Most ephemeral annuals do not have special drought-resistant or drought-tolerant adaptations: they are simply opportunists, much like the annuals of wetter communities which, however, tend to be successional rather than climax species. *See* PHOTOSYNTHESIS.

Succulent plants occur in all deserts but nowhere as conspicuously as in the Sonoran Desert of North America. These fleshy plants store water in their tissues, and tend to be especially prickly, a defense against browsers which need water as well as food. In American deserts the obvious succulents are Cactaceae; in Africa where cactus is almost absent, many species of Euphorbiaceae have the same types of growth form and occupy essentially the same ecological niches. *See* CACTUS; CARYOPHYLLALES.

Other perennial plants include trees and shrubs with long tap roots that can reach underground water, making them independent of scarce and unreliable rainfall. Some perennials burst into leaf only after rain, and are hence leafless most of the time. Yet others retain leaves throughout dry periods; the leaves tend to be small and narrow and require relatively little water to photosynthesize. These plants, mostly small shrubs and grasses, invariably have an extensive root system which enables them to maximize their acquisition of water. *See* PHYSIOLOGICAL ECOLOGY (PLANT); PLANT-WATER RELATIONS; PLANTS, LIFE FORMS OF.

Adaptations in animals. A majority of terrestrial invertebrates are cryptic in coloration, matching almost exactly components of the background on which they normally live. Many invertebrates and small vertebrates, such as rodents and snakes, spend the day in holes and burrows, and become active at night, when it is cooler. In all environments there are many small animals that escape by hiding, in most places to escape from predators; but in deserts escape from high daytime temperatures and radiation appears to be the chief adaptive response. The burrow constructed by the New World kangaroo rat, *Dipodomys*, protects the animal from daytime heat and radiation and also from predatory snakes and birds. The entrance of the burrow is sealed with soil which allows a high (30%) relative humidity to develop which helps the animal maintain its water budget. Kangaroo rats do not drink but exist on water derived from seeds they eat; their urine is the most concentrated of all mammals, another antidesiccation adaptation. Desert birds also shelter and avoid the heat of the day. In the Kalahari, the social weaverbird, *Philetarius socius*, builds a gigantic communal nest in an acacia or other suitable tree. The insulation in the nest protects the birds from the heat of the day and the cold of the night. *See* EXCRETION; PROTECTIVE COLORATION.

The largest of all birds, the ostrich, *Struthio camelus*, is an inhabitant of arid regions of Africa. The birds are too large to retreat into hiding during the day, although they may seek the shade of a big tree. With a high ambient temperature (95°F or 35°C or more) and no wind an ostrich loses heat by panting and by erecting the sparsely distributed feathers on its back. If there is wind, panting ceases and heat is lost by convection across the erect back feathers. At night when it is cool or even cold (below 64°F or 18°C), the back feathers are flattened and an insulating layer of air is trapped which reduces heat loss.

Soon after heavy rainfall, holes and depressions in sand or rock fill with water and within a few hours teem with microscopic life, chiefly algae, bacteria, and protozoa. After a few days countless small crustaceans appear, such as fairy and tadpole shrimps. These grow and reproduce with great rapidity. They can tolerate high water temperatures and also the high salinity which often builds up in temporary desert pools. The pools soon disappear and the land may then remain dry for months, even years, before the next downpour. The shrimps and other organisms diapause in the soil, usually as eggs which are remarkably resistant to high temperatures and extreme desiccation. *See* PHYSIOLOGICAL ECOLOGY (ANIMAL).

Aestivation. The term aestivation is used to describe a lengthy period of dormancy during which metabolism is much reduced. Many desert animals are capable of prolonged aestivation which enables them to survive during periods of food and water scarcity. In insects such as butterflies, which have a complete metamorphosis (clear-cut egg, larval, pupal, and adult stages), aestivation can occur at any stage of the life cycle but is almost invariably confined to a particular stage for a given species. Before entering aestivation, most animals seek out a secluded place that provides the best protection from temperature fluctuations and solar radiation, as well as from predators. Many species, ranging from lungfish to insect larvae, construct some form of protective cocoon in which to aestivate. *See* HIBERNATION AND ESTIVATION.

Migration and movement. An alternative to aestivation is migration or other movement away in search of better feeding and breeding conditions. Some of the African antelopes, such as the oryx, *Oryx gazella*, undertake long-distance movements which are correlated with changes in the quality and quantity of grass upon which they feed. These grass characteristics in turn are determined by rainfall, so there may be regular movements to and from areas where the rainfall is predictable and regular, or much more erratic movements in areas of unpredictable rainfall.

The semiregular migrations of several species of African weaverbirds (Ploceidae) are strongly associated with the seeding of wild grasses. The birds may form flocks of thousands or even tens of thousands of individuals and descend on and devour ripening seeds before moving on. Some of these birds, including the black-faced dioch, *Quelea quelea*, have become serious pests of millet and other cereals grown on land irrigated from underground water. *See* MIGRATORY BEHAVIOR.

The blooming of desert annuals and the leafing out of trees and shrubs soon after rainfall often result in a dramatic increase in abundance of leaf-eating insects. Eggs that have remained dormant for months produce larvae which can be so numerous as to defoliate vegetation. In Africa, white butterflies, *Belenois*, may migrate in vast numbers from areas where the larval food plants have been defoliated to areas where there is fresh growth. Pastoral people and their livestock undertake similar movements in response to changes in the availability of graze and browse for their animals, and to some extent, in response to changes in the risk of diseases such as trypanosomiasis, which affects both humans and domestic animals.

One of the most successful groups of desert animals is the grasshoppers. In many deserts there is a considerable variety of species, some of which are abundant. A few species are able to build up suddenly in numbers, often in response to unseasonal rain which has promoted a rapid growth of vegetation. They then undertake long-distance irruptive movements, followed by further breeding until they reach such numbers that they invade higher-rainfall areas and devastate crops. One of these grasshoppers, the desert locust, *Schistocerca gregaria*, can invade and seriously affect the vegetation of an area extending through 110° of longitude from West Africa to Assam, north to Turkey and south to southern Tanzania. Plagues of desert locusts may last for years and then suddenly stop, the locust reverting to a normal desert grasshopper.

Species diversity. The diversity of species of animals in a desert is generally correlated with the diversity of plant species, which to a considerable degree is correlated with the predictability and amount of rainfall. There is a rather weak latitudinal gradient of diversity with relatively more species nearer the Equator than at higher latitudes. This gradient is much more conspicuous in wetter ecosystems, such as forests, and in deserts appears to be overridden by the manifold effects of rainfall. Animals, too, may affect plant diversity: the burrowing activities of rodents create niches for plants which could not otherwise survive, and mound-building termites tend to concentrate decomposition and hence nutrients, which provide opportunities for plants to colonize.

Convergent evolution. Each desert has its own community of species, and these communities are repeated in different parts of the world. Very often the organisms that occupy similar niches in different deserts belong to unrelated taxa. The overall structural similarity between American cactus species and African euphorbias is an example of convergent evolution, in which separate and unrelated groups have evolved almost identical adaptations under similar environmental conditions in widely separated parts of the world. Convergent structural modification occurs in many organisms in all environments, but is especially noticeable in deserts where possibly the small number of ecological niches has necessitated greater specialization and restriction of way of life. The face and especially the large ears of desert foxes of the Sahara and of North America are remarkably similar, and there is an extraordinary resemblance between North American sidewinding rattlesnakes and Namib sidewinding adders. *See* EUPHORBIALES; ORGANIC EVOLUTION.

Desert community. Ecological change in deserts seems to occur slowly. Plant and animal succession is much less obvious than in other communities, although there is some evidence of cyclical change

in which species replace one another. In the Chihuahuan Desert of North America, the shrub *Larrea tridentata* reduces the wind speed enough to cause accumulation of soil and organic nutrients around the bottom of the trunk. This enables other plants to establish themselves, including the cactus *Opuntia leptocaulis*, whose seeds are introduced into the accumulated soil via the feces of rodents and birds which eat the fruit. The cactus grows and its roots overrun and outcompete those of the *Larrea* shrub, which then dies and falls over. Wind and water erosion then dislodge the rather shallow roots of the cactus, which itself dies and topples over. The vacant space is eventually filled by another *Larrea*, and the cycle starts again. No one has witnessed the complete cycle, which is believed to take many years to complete, but all stages can be identified and the presumed sequence of events drawn together. *See* ECOLOGICAL COMMUNITIES; ECOLOGICAL SUCCESSION; ECOLOGY; ECOSYSTEM.

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Desertification

Land degradation in low-rainfall and seasonally dry areas of the Earth. It can be viewed as both a process and the resulting condition. Desertification involves the impoverishment of vegetation and soil resources. Key characteristics include the degradation of natural vegetation cover and undesirable changes in the composition of forage species, deterioration in soil quality, decreasing water availability, and increased soil erosion from wind and water. Desertification is a global problem. Various stages of desertification can be seen in most of the world's drylands. In rare cases, desertification leads to abandoned, desertlike landscapes.

Causes and consequences. Although some authorities believe that climate change may be a causal factor, it is generally agreed that human activities, particularly excessive resource use and abusive land-use practices, are the primary cause of desertification. Specific activities leading to desertification include clearing and cultivation of low-rainfall areas where such cultivation is not sustainable, overgrazing of rangelands, clearing of woody plant species for fuelwood and building materials, and mismanagement of irrigated cropland leading to the buildup of mineral salts in the soil (salinization). Drought is often cited as a basic cause of desertification; however, it merely accelerates or accentuates land degradation processes already under way. *See* DROUGHT.

Consequences of desertification include reduced biological productivity, reduction of biodiversity, a gradual loss of agricultural potential and resource value, loss of food security, reduced carrying capacity for humans and livestock, increased risks from drought and flooding, and in extreme cases, barren lands that are effectively beyond restoration. Paleostudies, supported by model simulations, have shown that the intensity of Northern Hemisphere desert conditions has waxed and waned over the past 9000 years in response to the precession of the Earth's orbit about the Sun. Thus, it may be that the causal factors of desertification, whether climate change or human activities, depend on the time scale being addressed.

Concept. Desertification is an age-old problem in the world's drylands. For example, substantial areas in the region formerly known as Mesopotamia (the lower Tigris and Euphrates river basins of modern-day Iraq) have never recovered from desertification processes occurring over 4000 years ago. Likewise, much of the Mediterranean littoral has experienced varying degrees of desertification from excessive resource use beginning in the Roman era. Elsewhere in the world, on a lesser scale, desertification is known to have occurred centuries ago.

Beginning in the 1920s, there were warnings of the expansion or encroachment of the Sahara Desert into agrarian areas in the Sahel to the south. Massive soil erosion in the dust bowl in the United States during the 1930s pointed to the dangers of land degradation in drylands from unwise resource use. *See* SAHEL.

The term "desertification" was coined in 1949 by A. Aubreville, a French forester, who described the creation of desertlike conditions resulting from deforestation in the seasonally dry subhumid Sudanian belt south of the Sahel in Africa. It was not until the 1970s, however, that the term came into widespread international use. The reason was the drought in the African Sahel which began in 1968 and which by 1973 had resulted in at least 100,000 deaths and many times more refugees. It became obvious that the famine was caused by more than drought: an environmental calamity had been developing for many years due to unsustainable patterns of stockraising and cultivation that were incompatible with the region's environmental limitations. Drought served as a catalyst to reveal the damage caused by overgrazing and overcultivation. The resulting desertlike conditions made it seem that the Sahara was expanding. Similar land degradation was occurring in many other drylands, arousing fears of climate change and the specter of a global crisis.

As a result of these events, the United Nations (UN) convened the first international conference on desertification (UNCOD) in 1977. The conference aspired to provide a comprehensive understanding of desertification and its consequences for development, as well as to develop a plan of action to combat this global threat in its various settings. In 1978 the United Nations Environment Programme (UNEP) estimated that over 1.15×10^7 mi^2 (3×10^7 km^2) were severely desertified. In 1984 UNEP estimated that 35% of the world's land surface was at risk and that nearly 5.2×10^7 acres (2.1×10^7 hectares) of productive land were being reduced to near or complete uselessness each year. This agency

actively promoted the theory that the Sahara Desert was expanding southward at an average of almost 3.7 mi (6 km) per year. In response, a number of "green belt" schemes were implemented to try to halt the desert's expansion with rows of trees. In 1991 UNEP estimated that 73% of the world's rangeland, 47% of its rain-fed cropland, and 30% of its irrigated cropland was at least moderately desertified. This agency also asserted that the area irreversibly lost to desert was equivalent to all of the Earth's presently cultivated land and that desertification jeopardized the livelihoods of over a billion people in more than 100 countries. As a result of such claims, desertification became regarded as one of the world's most pressing environmental problems. During the 1992 UN "Earth Summit" Conference on Environment and Development in Rio de Janeiro, world leaders agreed to initiate negotiations on an international convention to address global desertification. The resulting Convention to Combat Desertification (CCD) was adopted in 1994. Following ratification by 50 countries, it became international law at the end of 1996. The convention's objective is to combat desertification and mitigate the effects of drought, particularly in Africa and other developing regions.

Controversies. Unfortunately, no universally accepted definition of desertification emerged from the 1977 UNCOD forum, and by the early 1980s there were more than a hundred different definitions of this phenomenon describing land degradation in ecosystems ranging from near-deserts to moist tropical rainforests. The lack of precision in defining the concept of desertification generated a series of controversies which gradually have called the validity of the term, and the concept itself, into question. It is argued that the term "desertification" should be avoided altogether in preference for the more general term "land degradation." The latter is preferred because it can be applied to all terrestrial ecosystems and avoids the misconceptions evoked by the term "desertification." There is also a basic disagreement over which ecosystems should be included in the purview of desertification research—drylands only, or all ecosystems in which degradation leads to drier conditions.

The struggle to define and determine the scope of desertification has become politicized. Primarily for political reasons—to focus international assistance efforts on poorer drought-prone developing countries, especially in the African Sahel—world leaders at the 1992 "Earth Summit" agreed to define desertification as "land degradation in arid, semi-arid, and dry subhumid areas resulting from various factors, including climatic variations and human activities." This definition was adopted by the CCD. However, the term "desertification" continues to be commonly applied in humid biomes as well—for example, for the degradation of tropical rainforest into savanna grassland.

There is also a fundamental controversy over the global extent of desertification. Most existing definitions of desertification are too general to be truly useful for estimating the extent or severity of this phenomenon. Global estimates are based on informed opinion rather than field research or remotely sensed data, and the extent of the problem may be grossly overstated. In the case of the Sahara Desert, the official wisdom during the late 1970s and 1980s was that the desert was expanding southward at a substantial rate of 3.4–5.6 mi (5.5–9 km) per year, but critics charged that this claim was based on a flawed 1975 study. Subsequent studies based on satellite image interpretation have conclusively demonstrated that there has been no permanent shift of the Sahara's southern boundary. These studies show that the Sahara boundary shifts from year to year in response to fluctuations in precipitation. According to these studies, rainfall fluctuations also adequately explain all variations in vegetative land cover from year to year. These studies seriously challenge the assertions by UNEP and other organizations that there has been a massive loss of land productivity in the Sahel from desertification. They also challenge estimates that tens of thousands of square miles of productive land are being lost to desertification each year.

There is a fundamental question about whether desertification—in the sense of extensive new desertlike environments being created through human activities and possibly climate change—is even occurring. By the mid-1990s, a revisionist critique of the conventional concept of desertification was firmly established. Key elements of this critique include the following: (1) Much of the data supporting the conventional concept of desertification was flawed, because it was collected during drought periods. (2) The concept relied too heavily on premature assumptions (for example, desert advance) and minimal field data. (3) It is exceedingly difficult to distinguish between the temporary effects of drought and the more permanent condition of desertification. This difficulty raises serious doubts about the validity of many traditional desertification studies and most estimates of the global extent of desertification. (4) Much of the literature on desertification ignores the fact that most of the world's drylands and desert areas are natural ecosystems. For example, the 35% of the Earth's land surface that is commonly claimed to be at risk of desertification includes the Sahara Desert and the world's other natural deserts—areas that are hardly at risk. (5) The traditional vision of desertification as the spread of desertlike conditions fails to recognize that land degradation in dryland areas often involves less obvious processes such as the gradual replacement of desirable forage plants by unpalatable species, with no net loss of primary productivity. (6) The imprecise and often-conflicting definitions of desertification have prevented a precise understanding of the actual problem and therefore have prevented effective treatment.

By the late-1990s, a major paradigm shift in drylands research was well established. Most specialists now acknowledge the difficulty of distinguishing between drought effects and human-induced degradation. Researchers increasingly focus on the regenerative capacity of dryland areas or on the complex localized interactions among climate variability, human

activities, and the land surface. Few specialists would deny that extensive degradation of vegetation and soil resources is occurring in many low-rainfall and seasonally dry environments. However, they increasingly concur that these are variations of land degradation processes also occurring in more humid environments, rather than a unique problem called desertification. *See* CLIMATE MODIFICATION; DESERT.

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Desiccant

A substance used to withdraw moisture from other materials. Although the removal of large quantities of water is done by evaporation, aided by moving air currents and by elevated temperature, the last traces of moisture are often held very tightly and do not evaporate readily. Furthermore, evaporation ceases when the moisture content of the material is reduced to that of the drying-air current. For final drying, a desiccant is used. This is a substance with a high affinity for water, that is, it is hygroscopic. It may retain water through capillarity or adsorption or by reacting chemically. A substance, such as calcium chloride, which retains enough water to become wet and dissolve is said to be deliquescent. The drying agent is placed directly into the gas or liquid to be dried; solid materials are placed in a desiccator, a closed vessel in which moisture escapes by vapor pressure and diffuses to the desiccant through the dry desiccator atmosphere. A desiccant loses potency as it takes on water; often it can be renewed by heating. Desiccants which form hydrates can be selected to maintain certain levels of low humidity in a closed vessel. *See* DELIQUESCENCE.

Allen L. Hanson

Important types. Silica gel possesses a high adsorptive power because of its extreme capillarity, the capillary pores occupying approximately 50% of its specific volume. The capillaries are probably spine-shaped, and the average pore diameter has been estimated to be 4×10^{-7} cm, which is only about 10 times the diameter of one molecule of adsorbate. The drying efficiency of silica gel depends upon the concentration of water in the gas mixture, the temperature of the gel and gas, the properties of the condensed liquid, its wettability, and the state of the gel itself. Silica gel adsorbs water vapor preferentially in the presence of other vapors. It is readily capable of drying air to a dew point that is below -70°C (-94°F). *See* DEW POINT.

Activated alumina is prepared from aluminum trihydrate. It is a granular porous adsorbent with properties similar to those of silica gel. Alumina gel has many applications in addition to gas drying. It is used to adsorb gases and vapors from gaseous mixtures and to dry liquids.

Anhydrous calcium sulfate is prepared from a high grade of gypsum, $CaSO_4 \cdot 2H_2O$, which is dried, crushed, sized, and heated to 230–260°C (450–500°F) for 2 h. This leaves a granular porous form of anhydrous calcium sulfate, with sufficient mechanical strength to support its own weight.

Magnesium perchlorate is the equal of any desiccant from the standpoint of drying efficiency. The adsorption rate is rapid, and the first hydrate does not lose water until 135°C (275°F) is reached, thereby permitting its use for drying gases at higher temperatures than most commercial desiccants. It passes through three hydrate stages, namely, di-, tri-, and hexahydrate. The last, $Mg(ClO_4)_2 \cdot 6H_2O$, represents saturation after adsorption of 48.6% of the dry weight of water, a high capacity. Although drying efficiency tends to decrease after each hydrate formation, even the trihydrate is superior to solid sodium hydroxide (NaOH) and calcium chloride ($CaCl_2$).

Other types. Other solid desiccants that have been used or studied for gas drying are oxides, such as barium and calcium oxide, and activated carbon. Barium oxide maintains a high drying activity up to 540°C (1000°F). Barium oxide also appears to have marked possibilities for the drying of gases at high temperatures.

Calcium oxide (CaO) has long been used as a desiccant because of its low cost. Although it has high drying efficiency, its capacity is low because of the formation of carbonates on its surface from the carbon dioxide of the air.

Activated carbon is old historically and probably has been as intensively studied as any single adsorbent. Although capable of adsorbing large amounts of water vapor, activated carbon finds its major use in solvent recovery, in odor and taste removal, and as a catalyst and catalyst carrier. Instead of the preferential adsorption of water vapor, as in the case of alumina and silica gel, organic vapors tend to displace any water present on the carbon. *See* ADSORPTION; DRYING.

William R. Marshall, Jr.

Design standards

Specifications of materials, physical measurement processes, performance of products, and characteristics of services rendered. Design standards may be established by individual manufacturers, trade associations, and national or international standards organizations. They may also be de facto standards; that is, they may be informally but universally accepted by those using them. The general purpose is to realize operational and manufacturing economies, to increase the interchangeability of products, and to promote uniformity of definitions of product characteristics.

Individual firms. Individual firms often maintain extensive and detailed standards of parts that are available for use in their product designs. Usually the standards have the effect of restricting the variety of parts to certain sizes and materials. For many firms, therefore, standardization has also meant simplification, defined as reducing a standard to the minimum of different sizes or other specifications that will cover user requirements. For instance, light bulbs and industrial electric motors have thus been limited to relatively few sizes. The practice has proved highly effective for many firms, especially when they had to reduce the "birthrate" of new-line items that threatened to clog even the most sophisticated inventory control system. A further advantage is that, provided the standards can be made sufficiently versatile, it is often possible to buy generic parts; for example, a certain bolt or screw can be bought as a standard hardware product rather than being a special job based on where it is installed.

A further development of this design approach may lead to the modulization of the entire product line, by reducing it to certain major subassemblies that are common to as many products as possible. Special jobs then typically require only a few added features, and cost savings may be realized.

A possible disadvantage is that the extensive use of general-purpose parts may jeopardize the spare parts business, especially where outside manufacturers can skim off the market for the more commonly used and profitable spare parts once the original patents, if any, have expired, and then leave the more complex and slow-moving spares to the original manufacturer. However, it is precisely this aspect of standardization which is often welcomed by the users of the product. For example, transfer machines, such as are used in automobile and other large-scale manufacture of piece parts, consist of many work heads joined by a conveyor, thus automating the materials-handling part of the process. Under pressure from the major auto makers, the mounting dimensions of the work heads have had to be standardized so that the entire machine can be reassembled out of the modules of several manufacturers instead of starting anew at each model change. Major manufacturers of diesel locomotives have likewise modulized and standardized their products, with great cost reductions relative to the former, highly individualistic steam locomotives, an action which helped displace them after 1945.

Standardization also determines the nature of design practice. Especially when the specifications also give data on strength and performance as well as the usual dimensions, it is only necessary to compute loads approximately and then select the nearest standard sizes. Much design effort is thereby saved, especially on detail drawings, bills of material, and so forth. This approach also simplifies programming when computer-aided design is used. *See* COMPUTER-AIDED DESIGN AND MANUFACTURING.

Trade associations. Trade associations are the principal sources of American industrial standards. These involve standardization over an entire product line. In general, their scope is considerably less than that within firms with extensive standardization programs, but the technical and policy considerations in the two levels of standardization are quite similar. Trade standards are primarily concerned with specifying overall dimensions, so that products of different manufacturers may be used interchangeably; with performance, so that customers know what they are buying; and with certain design features, such as major materials in order to assure proper function. In some cases, dimensional standards particularly must be related to standards in other industries; for instance, an American butter dish must accommodate the standard 4-oz (113-g) sticks in which butter is packed. Like national standards to which they are closely related, trade association standards should be established on the basis of as broad a consensus as possible within the industry. If standards were established such that any required burden of retooling and product change would fall in a discriminatory fashion upon only certain members of the industry, legal remedy would certainly be sought under the American antitrust laws.

In large part for this reason, a major problem in the development of standards by trade associations is the resistance to simplification which often arises. A line of least resistance then is simply to take everybody's prestandardization product lines, put them together, and call the result a standard. Historically, this has resulted in some standards having so many different products as scarcely to merit the name. For instance, there are dozens of steel compositions bearing standard designations but virtually identical to several others; in spite of these redundancies, they remain as standards.

In structural steel, some shapes were eliminated soon after the major standards were set about 1905, but simplification such as that in the equivalent British standards, for instance, never took place. Thus, the "nominal sizes" of steel beams are turned out in several weights per foot. However, economic analysis of weight-bearing capacity indicates that the lightest version of each size is usually best, except in the relatively unusual cases where there is a dimensional restriction such as a need for increased clearance.

National standards organizations. The principal industrial countries have official agencies that approve, consolidate, and in some cases establish standards. Among them are the British Standards Institution, German Institute for Norms, and the American National Standards Institute, which issue the BSI, DIN, and ANSI standards, respectively. ANSI is a federation of 900 companies, large and small, and about 200 trade, technical, professional, labor, and consumer organizations. It does not develop standards, but coordinates and promotes the voluntary development of national standards by industries provided that these have been established according to detailed rules for achieving a consensus among the producing industries, consumers (through ANSI's Consumer Council), relevant government agencies, and other interested parties.

There are about 10,000 approved ANSI standards, dealing with dimensions, ratings, terminology and symbols, test methods, and performance and safety specifications for equipment, components, and products. Major applications are in construction, electrical, electronic, mechanical, and nuclear products and processes, piping and welding, heating, air conditioning and refrigeration, information systems, medical devices, physical distribution, photography and motion pictures, and textiles. Most activities in the private and public sector use standardized products at some stage and specify them routinely in their purchases. ANSI standards have been embodied in building codes and many other government regulations, notably those of the Occupational Health and Safety Administration (OSHA). Related activities by government agencies also establish standards, as in agricultural grading.

A problem in setting standards in general, but particularly national ones, is the frequent need to anticipate future technologies and needs at an early stage of product development, knowing that later any substantial changes would be very difficult. For instance, domestic electrical supply is standardized at 110 V ac, 60 Hz, in the United States, and 210–250 V ac, 50 Hz, in Europe. These standards were set before 1900 and were, in part, the result of perceived limits on the mechanical capabilities of electrical machinery. Today, it would be better to have 400-Hz current such as used in aircraft systems because it would allow the use of smaller, faster, more efficient, and less expensive motors, as well as better functioning fluorescent lights. But any change would require costly conversions of transmission systems and consuming equipment.

Another example is that of railroad gages, which were set at 4 ft $8^1/_2$ in. (1435 mm) in most of Europe and North America and narrower in much of the rest of the world. Yet a wider gage of 7 ft (2 m) would allow for more efficient use and higher speeds, albeit at higher construction costs. Such changes have never been made, but others in established gages have often proved expensive, requiring new construction and inconvenient changes for passengers and transshipments of freight.

International standards. The national standardization agencies are members of a wide variety of international groupings and United Nations agencies. The principal ones are the International Organization of Standardization (ISO) and the International Electrotechnical Commission (IEC). These attempt to coordinate national activities and promote cooperation in the area of standardization. Several of the more than 50 organizations deal with weights and measures, which have always been the subject of extensive international coordination and agreement. Others engage in transnational or international activities in the standardization of many products or cover specific regional issues and requirements. Some, like the European Economic Community (EEC), the International Telecommunications Union (ITU), the International Civil Aviation Organization (ICAO), or the administration of the General Agreement on Tariffs and Trade (GATT) mainly have other political, economic, and scientific concerns but must necessarily take note of standardization as part of their work. *See* ENGINEERING; ENGINEERING AND ARCHITECTURAL CONTRACTS; ENGINEERING DESIGN; PHYSICAL MEASUREMENT; SCIENTIFIC METHODS; STANDARD. John E. Ullman

Desmodorida

An order of nematodes in which the variable stoma may or may not be armed with a dorsal tooth which may or may not be opposed by subventral denticles. The variable amphids range from reniform-to-elongate loops to simple or multiple spirals. The cephalic sensilla are generally in three whorls; however, the second and third whorls may be combined. Distinguishing the order is the cephalic capsule or helmet and the conspicuous somatic annuli. In some groups anterior and posterior adhesion tubes are utilized in locomotion.

There are five desmodorid superfamilies: Ceramonematoidea, Desmodoroidea, Draconematoidea, Epsilonematoidea, and Monoposthioidea. Ceramonematoidea comprises marine forms with distinctive cuticular ornamentation that takes the form of crested annuli. Nothing is known of the feeding habits. Most species of Desmodoroidea are found in the marine environment. Draconematoidea is an unusual superfamily of marine forms; when relaxed, the body is dorsally and then ventrally arched into a shallow sigmoid shape. Epsilonematoidea and Monoposthioidea both comprise marine forms. *See* NEMATA. Armand R. Maggenti

Desmoscolecida

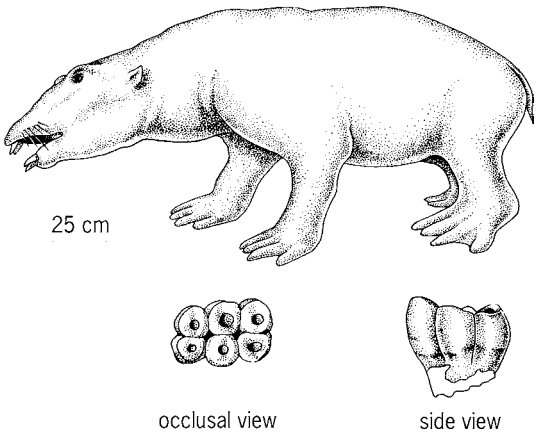
An order of nematodes in which the taxa are readily recognized by their conspicuous body annulation. The annuli may be covered with concretion rings, or the cuticle may be ornamented with scales, warts, or bristles. The cephalic sensilla reportedly are reduced in number. The internal whorl is absent, the second whorl is papilliform, and the four sensilla of the third circlet are setiform. The vesiculate amphids are oval to circular and occupy much of the cephalic region. The somatic setae are tubular, and the open distal ends are often elaborate. The posterior esophagus is only slightly expanded. When pigment spots or ocelli are present, they are just posterior to the esophagus. Females are amphidelphic, and the ovaries are generally outstretched. The male spicules are generally accompanied by a gubernaculum. Both sexes have three caudal glands.

There are two desmoscolecid superfamilies, Desmoscolecoidea and Greeffielloidea. Most species of Desmoscolecoidea are found in the marine environment, although some species have also been collected from brackish waters. Greeffielloidea are free-living nematodes found primarily in marine or, rarely, estuarine habitats. *See* NEMATA. Armand R. Maggenti

Desmostylia

An extinct order of large, hippopotamuslike, amphibious, gravigrade, shellfish-eating mammals, nearly 12 ft (4 m) long, that frequented shallow bays and coastal margins during the Oligocene and Miocene. Their main distribution was northern trans-Pacific, but discoveries in Florida indicate a much broader distribution.

There are two families, the Desmotylidae (*Desmostylus*, *Cornwallius*, and *Vanderhoofia*) and the Paleoparadoxidae (*Paleoparadoxia*). The many cusped molars are composed of a cluster of heavy enameled cylindrical columns (see **illus.**), resembling the pavement teeth in drumfish. The cheek teeth are antero-vertically replaced as in the Proboscidea and Sirenia. There are one to four pairs of procumbent tusks. These animals probably descended from an ancestral stock that also gave rise to the Proboscidea and Sirenia. *See* MAMMALIA; PROBOSCIDEA; SIRENIA.



A cheek tooth and restoration of *Desmostylus* from the Miocene of California.

Gideon T. James

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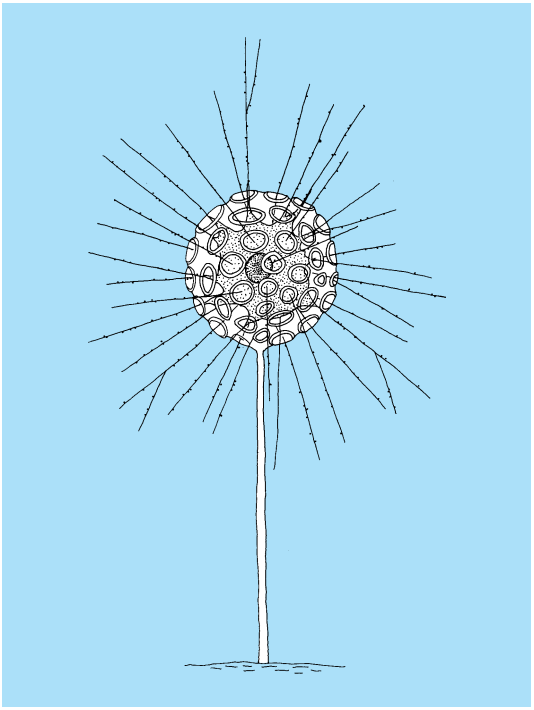
Desmothoracida

An order of Heliozoia. These ameboid organisms live in a stalked perforated organic or inorganic lorica. Stiffened pseudopodia extend from the perforations to capture food which collides with them. The best-known species is *Clathrulina elegans,* found in fresh-water environments (see **illus.**). This species has a spherical organic test. Unlike most Heliozoia it has a life cycle which involves motile flagellated cells which arise from the sessile form and then may settle, transforming into amebas with thin pseudopodia. These ameboid forms secrete the stalk and lorica to give rise to the typical sessile (trophic) organism. Under unfavorable conditions the organism encysts within the lorica. *See* HELIOZOIA; PROTOZOA; SARCODINA.



Sessile form of *Clathrulina elegans*. (*After C. F. Bardele, Cell cycle, morphogenesis and ultrastructures in the pseudoheliozoian Clathrulina elegans, Z. Zellforsch., 150:219–242, 1972*)

David J. Patterson

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Desorption

A process in which atomic and molecular species residing on the surface of a solid leave the surface and enter the surrounding gas or vacuum. In stimulated desorption studies, species residing on a surface are made to desorb by incident electrons or photons. Measurements of these species provide insight into the ways that radiation affects matter, and are useful analytical probes of surface physics and chemistry. In thermal desorption studies, adsorbed surface species are caused to desorb as the sample is heated under controlled conditions. These measurements can provide information on surface-bond energies, the species present on the surface and their coverage, the order of the desorption process, and the number of bonding states or sites.

Stimulated desorption. Stimulated desorption from surfaces is initiated by electronic excitation of the surface bond by incident electrons or photons. The classical model of desorption is an adaptation of the theory of gas-phase dissociation, in which desorption results from excitation from a bonding state to an antibonding state. In this model the excitations

for gas-phase dissociation and desorption from surfaces are identical, but the surface, which can absorb energy from the excited adsorbate before it can desorb, dramatically reduces the overall dissociation yield from surfaces. Thus, processes which are strong in the gas phase can be quenched on the surface, leading to a marked difference in the relative importance of competing mechanisms in the two environments. This model applies predominantly to neutral desorption of adsorbed molecular species.

Another model, proposed by M. L. Knotek and P. J. Feibelman, which is more applicable to the phenomenon of ion desorption, was first observed in studies of the desorption of positively ionized oxygen (O^+) from the surface of titanium(IV) oxide (TiO_2). Here it is found that O^+ is desorbed not by valence level excitation, but by ionization of the titanium and oxygen core levels. These levels, of course, have little to do with bonding. Furthermore, the fact that the oxygen is desorbed as an O^+ ion (whereas it is nominally at O^{2-} on the surface) implies a large (three-electron) charge-transfer preceding desorption.

Titanium(IV) oxide is an ionic solid, and more importantly it is a maximal valency ionic compound. Maximal valency means that in the solid the titanium (Ti) is ionized to the noble-gas configuration Ti^{4+}, so that there are effectively no valence electrons on the Ti ion in the solid; its highest occupied orbital is $3p$ at about 30 eV below the Fermi level. The important thing about maximal valency ionic materials is that if a core level is ionized in either the cation or the anion, higher-lying valence electrons will decay into the deep hole left in the core level and give up the energy gained by emitting multiple-valence electrons from the solid (multiple Auger emission). Since valence electrons are only on the anion, the anion is selectively stripped in the process. In TiO_2 one electron is lost from O^{2-} by decay into the core hole; in about 10% of the events two Auger electrons are emitted, transforming the O^{2-} to an O^+ ion. An O^+ ion in a highly repulsive potential for positive ions results, which leads to O^+ desorption. In the general case of an ionic bond M^+A^- (**Fig. 1**), the incident radiation ionizes the core level of the M^+A^- pair. Auger decay of the core hole selectively removes charge from the A^- ion and results in its being transformed into an A^+, and the resulting repulsion causes the A^+ to desorb.

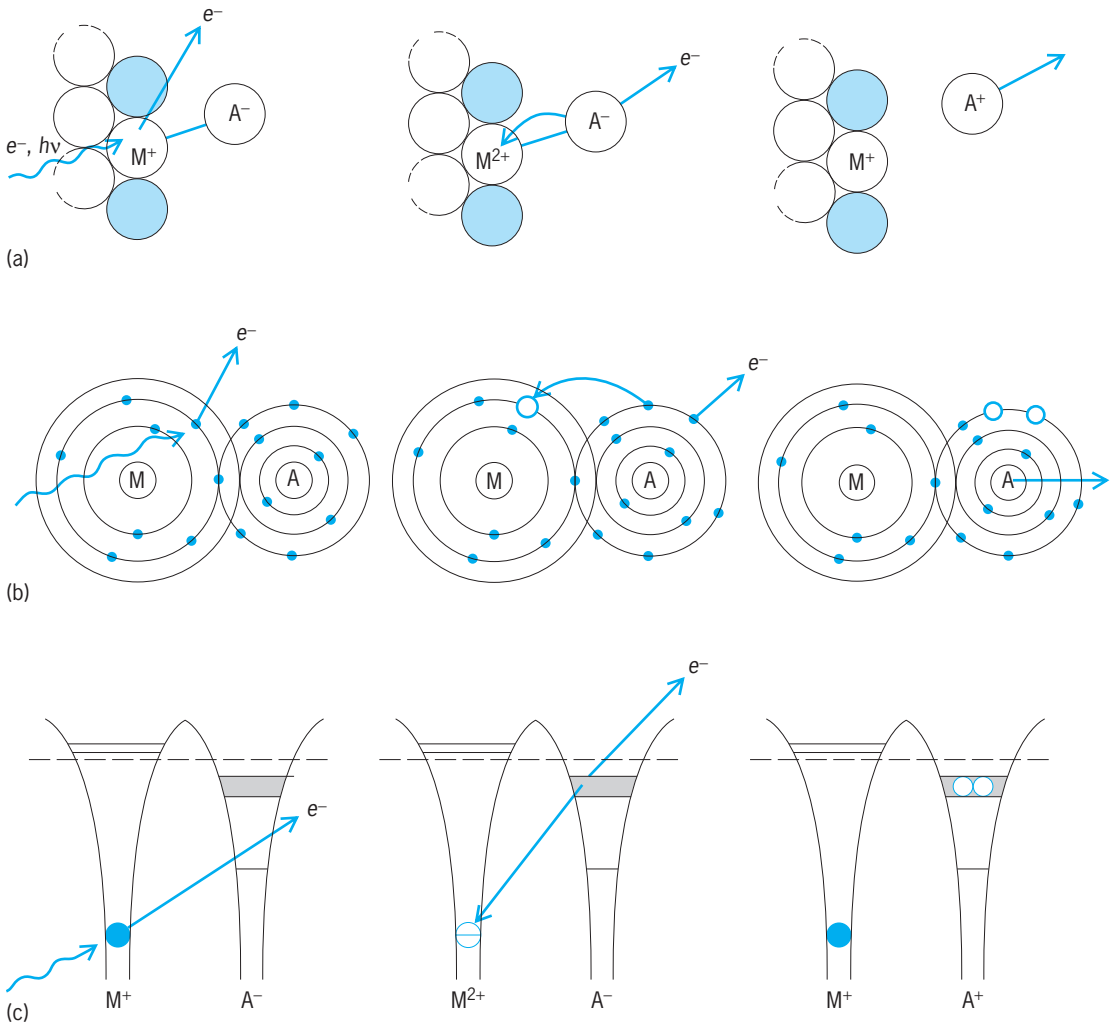

Fig. 1. Sequence of events leading to stimulated desorption by an electron (e^-) or photon ($h\nu$): (*a*) in the structural model, (*b*) in terms of orbitals, and (*c*) in terms of energy levels.

This model applies to a wide range of ionic materials, and explains why maximal valency materials like TiO_2, vanadium(V) oxide (V_2O_5), tungsten(VI) oxide (WO_3), molybdenum(VI) oxide (MoO_3), sodium fluoride (NaF), lithium fluoride (LiF), sodium chloride (NaCl), and so forth, are decomposed by irradiation. *See* AUGER EFFECT.

This mechanism for desorption can also be effective for covalently bonded surface species. Covalently bonded structures differ from ionically bonded structures in that the electronic charge which bonds the atoms together is distributed more evenly through the structure. The electronic interaction between atoms is strong so that an electronic excitation produced on one atom can quickly move to a neighbor. Thus, a simple electronic excitation can quickly move off the site where it was created before dissociation can occur. It typically takes 10^{-12} s for bond dissociation to occur, whereas the motion of simple electronic excitations can occur in 10^{-16} s. However, an important attribute of states where two or more electrons are removed from the atom is that such multiply excited states are much more difficult to move to a neighbor and hence are localized for times long enough for desorption to occur. Several features of a covalent bond can enhance this tendency to localize on the atom—for example, reducing the number of neighbors (and hence bonds), or reducing the bond's ability to transport charge to the desorbing species to neutralize the excited state.

Applications. Stimulated desorption studies are finding wide use. First, they can show the ways in which radiation affects the structure of solids. This will have important applications in the areas of radiation-induced damage and chemistry. Second, as an analytical tool, they offer a unique new way to study the physics and chemistry of atoms on surfaces which, when combined with the many other surface techniques based largely on electron spectroscopy, can provide new insight. Finally, models of the surface bond are put to a much sterner test in attempting to explain desorption phenomena. *See* RADIATION DAMAGE TO MATERIALS.

Analysis of surface structure. The techniques of electron- and photon-stimulated desorption can be very useful in the study of how atoms and molecules bond to surfaces. Since ions are desorbed only from the topmost layers of the solid, these techniques study only that outermost layer. By mass-analyzing the desorbed atomic and molecular species, the nature of the chemical species on the surface can be deduced. More importantly, an adsorbed species will be desorbed when the core level of its bonding-site atom is ionized. Since the energy of the electron or photon necessary to ionize a given atom's core level is well known and characteristic of that atom, the bonding sites of each of the desorbing species can be determined. Thus surface-, site-, and adsorbate-specific information can be obtained. Elaborate methods for deducing how the bonding site and adsorbate atoms are electronically and structurally configured can be utilized with this technique. The method is almost the only spectroscopy capable of detecting hydrogen, a classic problem in this field.

Ion angular distributions. An additional important discovery is that ion angular distributions from stimulated desorption are not isotropic, but show that ions are emitted in relatively narrow cones which project along the nominal ground-state bond directions. An angular display shows an azimuthal pattern which reflects the symmetry of the bonding site, and a polar angular distribution which reflects the bonding angle relative to the surface normal. Thus this technique provides a direct display of the surface-bonding geometry.

There are a number of ways to measure the angular distributions of desorbed ions. One of the simplest employs an image intensifier. Individual ion events result in light pulses at a phosphor screen, and time-averaged photographs of the screen provide visual images of the distributions and thus allow a direct view of the arrangement of atoms and molecules on the surface (**Fig. 2**). The directness and simplicity of the technique make it particularly attractive to measure complicated surface adsorbate structures or chemical configurations which do not contain the long-range order necessary in many other techniques for measuring structure. *See* LIGHT AMPLIFIER.

Thermal desorption. Thermal desorption mass spectroscopy is possibly the oldest technique for the study of adsorbates on surfaces. Three primary forms of the thermal desorption experiment involve measurement of (1) the rate of desorption from a surface during controlled heating (temperature-

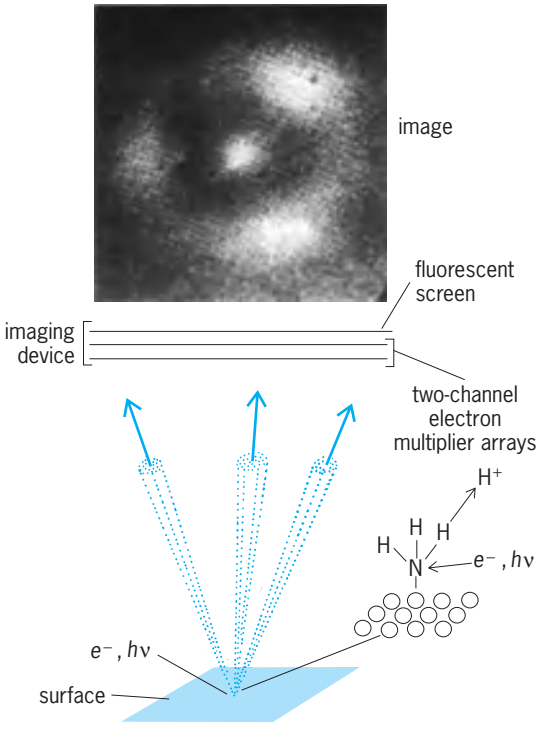

Fig. 2. Measurement of the angular distribution of ions leaving a surface in stimulated desorption.

programmed thermal desorption), (2) the rate of desorption at constant temperature (isothermal desorption), and (3) surface lifetimes and diffusion under exposure to a pulsed beam of adsorbates (molecular-beam experiments). Of the three, temperature-programmed thermal desorption is by far the most widely applied, and the discussion below is confined to it. *See* MOLECULAR BEAMS.

In thermal desorption measurements, a surface is first dosed with an adsorbate gas, and then the sample is heated at a controlled rate β while the rate of desorption of the adsorbed gas is measured with a mass-filter gas analyzer. The most straightforward information provided is the nature of the desorbed species from mass analysis, and a determination of the absolute coverage by the adsorbate, which is very difficult to obtain with other techniques. Coverage is determined by measuring the total amount of gas desorbed. The technique can also provide several important kinetic parameters of the desorption process as follows.

Since adsorption reactions are usually nonactivated, the desorption activation energy E, which thermal desorption can provide, is a good approximation to the differential heat of adsorption. Thus these studies provide information on surface-bond energies. *See* ADSORPTION.

Desorption kinetics can be quite complex, especially when there are strong interactions between adsorbates. The general form of the mth-order desorption rate equation is given by Eq. (1), where $R(t)$

$$R(t) = \nu_0^{(m)} n^m \exp(-E/kT_s) \qquad (1)$$

is the rate of desorption as a function of time, T_s is the substrate temperature, $\nu_0^{(m)}$ is the preexponential factor of the desorption-rate coefficient for order m, and n is the two-dimensional adsorbate concentration. Desorption has been observed to obey zero-, first-, and second-order kinetics. Zero-order kinetics is observed for multilayers where removal of an adsorbed species does not lower the effective coverage feeding the desorption process. First-order kinetics is observed for simple desorption of an atomic or molecular species from a submonolayer of adsorbate. Second-order kinetics is often observed in systems where dissociative adsorption has occurred and recombination reactions occur before the desorption.

Desorption spectra can determine the number of binding states of an adsorbate on a surface, from the number of desorption peaks observed, and, coupled with structural models of the surface, can help determine both the nature of the bonding sites and the bonding geometries.

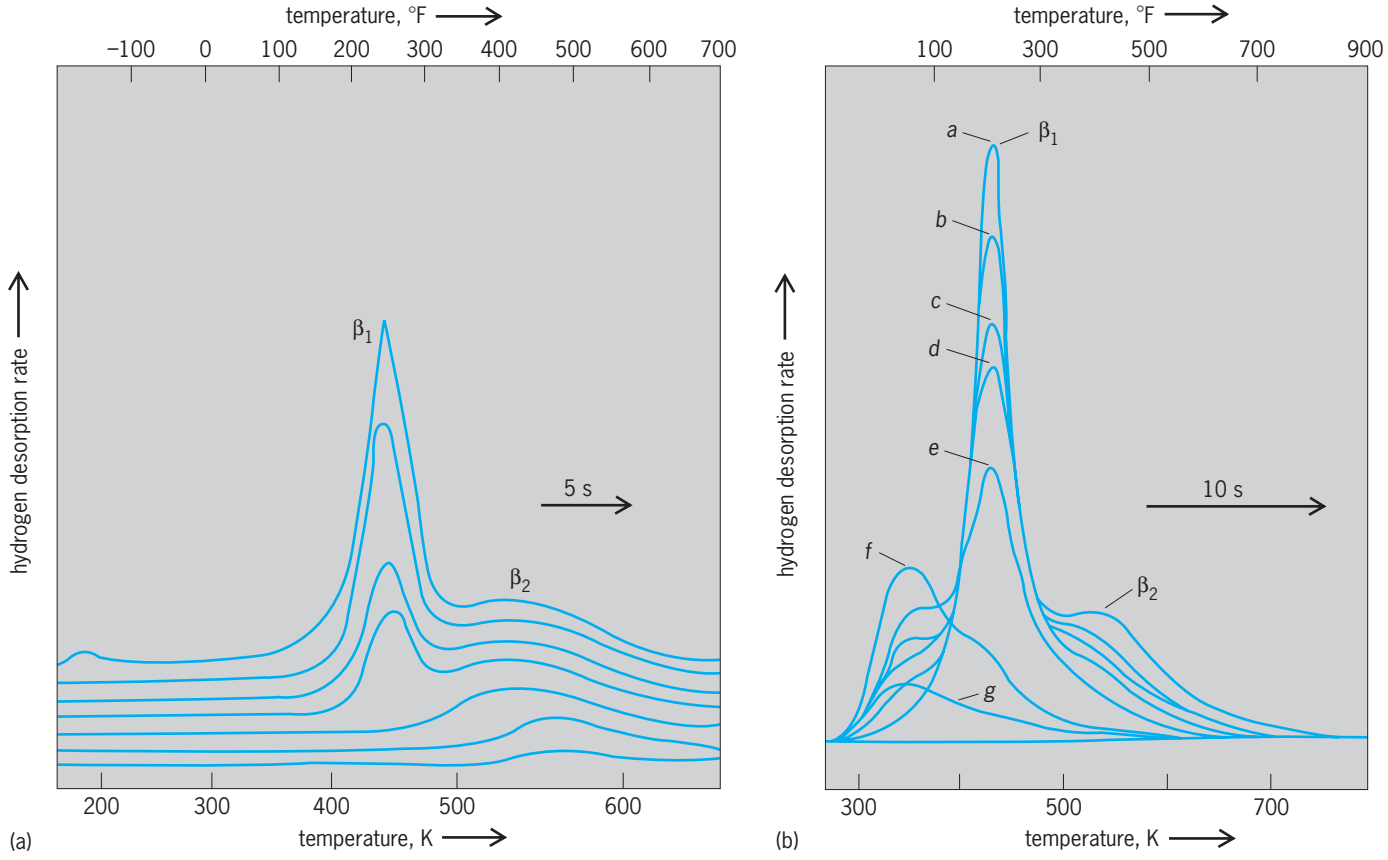

Fig. 3. Hydrogen desorption rate as a function of sample temperature for a heating rate of approximately 10°C/s. (*a*) Successive curves from bottom to top represent increasing hydrogen coverage. (*b*) Desorption when carbon monoxide (CO) is added to the hydrogen-preadsorbed surface. Successive curves from *a* to *g* represent increasing carbon monoxide coverage. (*After T. E. Madey and J. T. Yates, Desorption method A: Probes of kinetics and bonding at surfaces, Surf. Sci., 63:203–231, 1977*)

The preexponential factor of the desorption-rate coefficient, equivalent to an attempt frequency, the rate at which the desorbing particle attempts to leave the surface, not only is necessary to understand desorption kinetics, but helps in understanding the general problem of surface chemical dynamics.

The existence of indirect and direct interactions between coadsorbed species can be determined from thermal desorption. Knowledge of such interactions is directly applicable to such important problems as catalysis and corrosion. *See* CATALYSIS; CORROSION.

There are several methods to derive the kinetic parameters of Eq. (1) experimentally. If a value is assumed for the preexponential factor ν_0 (usually of the order of 10^{13}), the desorption activation energy is given by Eq. (2) for first-order kinetics, and by Eq. (3)

$$E/kT_p = (\nu_0^{(1)}/\beta)\exp(-E/kT_p) \tag{2}$$

$$E_k T_p^2 = (\nu_0^{(2)}\, n_0/\beta)\exp(-E/kT_p) \tag{3}$$

for second-order kinetics. Here β is the heating rate, n_0 is the initial adsorbate concentration, and T_p is the temperature of peak desorption rate. Alternatively, β can be varied and E can be derived from Eq. (4) for

$$E/kT_p = d(\ln\beta)/d(\ln T_p) \tag{4}$$

first order; for second order, E/k is determined from the slope of a plot of $\ln(n_0 T_p^2)$ versus $1/T_p$. It is also possible to derive the desorption energy from the width of the desorption peak.

One of the simplest of surface systems, hydrogen adsorbed on the (100) surface of tungsten (W), provides graphic examples of how the above types of information manifest themselves and are analyzed. **Figure 3** shows molecular desorption of hydrogen as a function of sample temperature for heating rates β as shown in the figure. At low hydrogen coverages a single β_2 state is desorbed with second-order kinetics. At coverages above 0.2 monolayer of hydrogen, the β_1 state desorbs with first-order kinetics. When carbon monoxide (CO) is added to the preadsorbed hydrogen layer, it causes the depopulation of both the β_1 and β_2 states and introduces a weak state of lower binding energy. The β_2 state is now known to be hydrogen bound at reconstructed sites on the W(100) surface, while the β_1 state is hydrogen at unreconstructed sites. *See* CRYSTAL STRUCTURE.

While the thermal desorption techniques are among the simplest of surface probes, they remain indispensable because of their directness and the variety of information they convey. Thus while surface science moves to detailed methods involving extremely sophisticated apparatus, the simple thermal desorption methods remain an important part of the overall picture. *See* SURFACE PHYSICS. M. L. Knotek

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Destructive distillation

The primary chemical processing of materials such as wood, coal, oil shale, and some residual oils from refining of petroleum. It consists in heating material in an inert atmosphere at a temperature high enough for chemical decomposition. The principal products are (1) gases containing carbon monoxide, hydrogen, hydrogen sulfide, and ammonia, (2) oils, and (3) water solutions of organic acids, alcohols, and ammonium salts. For a discussion of the products of the destructive distillation, or coking, of coal *see* COAL CHEMICALS

Crude shale oil may be subjected to a destructive, or coking, distillation to reduce its viscosity and increase its hydrogen content. Subsequent catalytic hydrogenation (cobalt molybdate alumina catalyst, about 750°F or 400°C, 100–1500 lb/in.^{2} or 700–10,000 kilopascals pressure) lowers the nitrogen and sulfur contents so that the oil can then be refined by normal petroleum refinery operations. Residual oils from petroleum refinery operations are subjected to coking-distillation to reduce the carbon content. The coke is used for the manufacture of electrode carbon and the oil is returned to the feed for normal petroleum refining.

Prior to about 1920, destructive distillation of wood was an important source of methanol, acetic acid, and acetone. These chemicals are now produced at a lower cost by other methods; for example, methanol is produced by hydrogenation of carbon monoxide, acetaldehyde and acetic acid from acetylene, and acetone by fermentation processes and by decomposition of cumene hydroperoxide. The main product of the destructive distillation of wood is 40–45% charcoal used in metallurgical processes in which the low content of ash, sulfur, and phosphorus is important. Special chars made by the destructive distillation of the shells of nuts have very large surface areas and are used as adsorbents in gas masks and in some chemical processes. Char from the destructive distillation of bones is used for decolorizing solutions of raw sugar, and the oil obtained during the decomposition of bones contains recoverable amounts of pyridine, pyrrole, quinolines, and other nitrogen compounds. *See* CHARCOAL; COKE; COKING (PETROLEUM); OIL SHALE; PYROLYSIS; WOOD CHEMICALS. H. H. Storch; H. W. Wainwright

Detergent

A substance used to enhance the cleansing action of water. A detergent is an emulsifier, which penetrates and breaks up the oily film that binds dirt particles, and a wetting agent, which helps them to float off. Emulsifier molecules have an oillike nonpolar

portion which is drawn into the oil, and a polar group that is water-soluble; by bridging the oil-water interface, they break the oil into dispersible droplets (emulsion). As a surfactant, a detergent decreases the surface tension of water to help it penetrate soil. *See* EMULSION.

Soap, the sodium salt of long-chain acids, was the principal detergent until superseded in 1954 by synthetic detergents (syndets) which, unlike soap, do not form insoluble products with the calcium in hard water. Most syndets are of the anionic type, that is, sodium salts of alkyl sulfates or sulfonates which contain a chain of 7–18 carbon atoms; an example is sodium lauryl sulfate, $CH_3(CH_2)_{11}OSO_3^-,Na^+$. Alkyl benzene sulfonates (ABS) with branched carbon chains were found to persist in wastewater and have been replaced by linear alkyl benzene sulfonates (LAS), which are biodegradable by bacterial action. Anionic detergents are best for water-absorbing fibers such as cotton, wool, and silk. Nonionic detergents are polyethers made by combining ethylene oxide with a 12-carbon lauryl alcohol obtained from tall oil, a waste product from the paper industry. They are used for water-repelling "permanent press" fabrics, and their low-foaming property is desirable for automatic washers. Cationic syndets are quaternary base compounds of type $(NR_4)^+,X^-$. They are more expensive, but some are germicidal; some are used as fabric softeners and as good metal cleaners. *See* SOAP.

Detergents must contain alkaline "builders" to bind dissolved metal ions and support emulsification. Sodium pyrophosphate or polyphosphate is preferred because of low cost and high cleansing effectiveness. However, when discharged with laundry wastewater, these compounds supply nutrient to phosphate-deficient lakes and streams and thus lead to eutrophication, and their use is now banned by law. For a brief time an effective builder designated NTA was promoted as a phosphate substitute, until it was found that it solubilizes metal and may lead to metal poisoning. Less harmful, but less effective, builders such as sodium carbonate are now widely used in detergents. *See* EUTROPHICATION.

Many additives are used in detergents to provide scent, brightening (usually through fluorescent action), or bleaching action. For a time, enzymes which cleave protein molecules (proteolytic) or fats (lipolytic) were widely used as additives, but they do not add significantly to cleansing action. Biodegradability is essential for detergents; it ensures that components of detergents will be broken down by bacterial action before undesirable aftereffects can occur. Nonbiodegradable detergents can prevent effective bacterial action in septic tanks and sewage treatment plants, and can cause undesirable persistent foaming in rivers.

Allen L. Hanson

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Determinant

The concept of a determinant can best be explained with reference to a matrix A. Let

$$A = \begin{pmatrix} a_{11} & a_{12} & \cdots & a_{1n} \\ a_{21} & a_{22} & \cdots & a_{2n} \\ \cdots & \cdots & \cdots & \cdots \\ a_{n1} & a_{n2} & \cdots & a_{nn} \end{pmatrix}$$

be an n by n square array of numbers. Such an array is called an n by n square matrix, and the numbers a_{ij} in the array are called elements of the matrix. Determinants are valuable in the solution of sets of linear equations. The determinant of the matrix A, as indicated by Eq. (1), is a number which

$$|A| = \begin{vmatrix} a_{11} & a_{12} & \cdots & a_{1n} \\ a_{21} & a_{22} & \cdots & a_{2n} \\ \cdots & \cdots & \cdots & \cdots \\ a_{n1} & a_{n2} & \cdots & a_{nn} \end{vmatrix} \tag{1}$$

is the value of a certain function of the elements of A.

Before defining the determinant $|A|$, it may be noted that the theory of determinants had its origin in the solution of linear systems of equations and that determinants occur in virtually all branches of mathematics and related fields. *See* LINEAR SYSTEMS OF EQUATIONS; MATRIX THEORY; POLYNOMIAL SYSTEMS OF EQUATIONS.

In the linear system of equations shown as Eqs. (2), successive elimination of the unknowns x and y yields a simpler equivalent system in Eqs. (3). The

$$\begin{aligned} a_{11}x + a_{12}y &= b_1 \\ a_{21}x + a_{22}y &= b_2 \end{aligned} \tag{2}$$

$$\begin{aligned} (a_{11}a_{22} - a_{12}a_{21})x &= b_1a_{22} - a_{12}b_2 \\ (a_{11}a_{22} - a_{12}a_{21})y &= a_{11}b_2 - b_1a_{21} \end{aligned} \tag{3}$$

coefficients of Eqs. (3) are determinants of 2 by 2 matrices

$$\begin{vmatrix} a_{11} & a_{12} \\ a_{21} & a_{22} \end{vmatrix} = a_{11}a_{22} - a_{12}a_{21}$$

$$\begin{vmatrix} b_1 & a_{12} \\ b_2 & a_{22} \end{vmatrix} = b_1a_{22} - a_{12}b_2$$

$$\begin{vmatrix} a_{11} & b_1 \\ a_{2;} & b_2 \end{vmatrix} = a_{11}b_2 - b_1a_{21}$$

with elements which are the coefficients of Eq. (2).

If

$$\begin{vmatrix} a_{11} & a_{12} \\ a_{21} & a_{22} \end{vmatrix} \neq 0$$

then the solution of Eqs. (2) is readily given by

ratios of the above determinants. Similarly in the system

$$a_{11}x + a_{12}y + a_{13}z = b_1$$

$$a_{21}x + a_{22}y + a_{23}z = b_2$$

$$a_{31}x + a_{32}y + a_{33}z = b_3$$

the determinant shown in Eq. (4) enters into the solution of the system.

$$\begin{vmatrix} a_{11} & a_{12} & a_{13} \\ a_{21} & a_{22} & a_{23} \\ a_{31} & a_{32} & a_{33} \end{vmatrix} = a_{11}a_{22}a_{33} + a_{12}a_{23}a_{31} + a_{13}a_{21}a_{32} - a_{13}a_{22}a_{31} - a_{12}a_{21}a_{33} - a_{11}a_{23}a_{32} \quad (4)$$

The determinant of Eq. (1) is called a determinant of order n, and it can be defined by induction on n. For explanatory purposes, determinants of orders $n = 1$, 2, and 3 are defined as follows:

For $n = 1$,

$$|a_{11}| = a_{11}$$

For $n = 2$,

$$\begin{vmatrix} a_{11} & a_{12} \\ a_{21} & a_{22} \end{vmatrix} = a_{11}\,|a_{22}| - a_{12}\,|a_{21}| = a_{11}a_{22} - a_{12}a_{21}$$

For $n = 3$,

$$\begin{vmatrix} a_{11} & a_{12} & a_{13} \\ a_{21} & a_{22} & a_{23} \\ a_{31} & a_{32} & a_{33} \end{vmatrix}$$

$$= a_{11}\begin{vmatrix} a_{22} & a_{23} \\ a_{32} & a_{33} \end{vmatrix} - a_{12}\begin{vmatrix} a_{21} & a_{23} \\ a_{31} & a_{33} \end{vmatrix} + a_{13}\begin{vmatrix} a_{21} & a_{22} \\ a_{31} & a_{32} \end{vmatrix}$$

$$= a_{11}(a_{22}a_{33} - a_{23}a_{32}) - a_{12}(a_{21}a_{33} - a_{23}a_{31}) + a_{13}(a_{21}a_{32} - a_{22}a_{31})$$

This definition is seen to agree with the definitions for determinants of order 2 and 3 and, by the principle of mathematical induction, defines a determinant of order n for every n. There are many other definitions of a determinant of order n which can be proved to be equivalent to the definition given here. The determinants of order 2 and 3 are defined in terms of determinants of order 1 and 2, respectively. Suppose now that a determinant of order $n - 1$ has been defined. In Eq. (1), denote by A_{1j}, $j = 1, 2, \ldots, n$, the determinant of order $n - 1$ formed by deleting from Eq. (1) the first row and the jth column. Equation (5) then, by definition, holds.

$$|A| = a_{11}A_{11} - a_{12}A_{12} + a_{13}A_{13} + \cdots + (-1)^{n-2}a_{1n-1}A_{1n-1} + (-1)^{n-1}a_{1n}A_{1n} \quad (5)$$

As an example of Eq. (5), the relations shown in Eqs. (6) are formed.

$$|A| = \begin{vmatrix} -2 & 0 & 1 & 3 \\ 5 & 2 & -6 & 1 \\ 0 & 4 & -1 & 7 \\ 10 & 2 & 3 & 1 \end{vmatrix}$$

$$= (-2)\begin{vmatrix} 2 & -6 & 1 \\ 4 & -1 & 7 \\ 2 & 3 & 1 \end{vmatrix} - (0)\begin{vmatrix} 5 & -6 & 1 \\ 0 & -1 & 7 \\ 10 & 3 & 1 \end{vmatrix} + (1)\begin{vmatrix} 5 & 2 & 1 \\ 0 & 4 & 7 \\ 10 & 2 & 1 \end{vmatrix} - (3)\begin{vmatrix} 5 & 2 & -6 \\ 0 & 4 & -1 \\ 10 & 2 & 3 \end{vmatrix} \quad (6a)$$

$$\begin{vmatrix} 2 & -6 & 1 \\ 4 & -1 & 7 \\ 2 & 3 & 1 \end{vmatrix}$$

$$= (2)\begin{vmatrix} -1 & 7 \\ 3 & 1 \end{vmatrix} - (-6)\begin{vmatrix} 4 & 7 \\ 2 & 1 \end{vmatrix} + (1)\begin{vmatrix} 4 & -1 \\ 2 & 3 \end{vmatrix}$$

$$= 2(-1 - 21) + 6(4 - 14) + (12 + 2) = -90 \quad (6b)$$

$$\begin{vmatrix} 5 & 2 & 1 \\ 0 & 4 & 7 \\ 10 & 2 & 1 \end{vmatrix}$$

$$= (5)\begin{vmatrix} 4 & 7 \\ 2 & 1 \end{vmatrix} - (2)\begin{vmatrix} 0 & 7 \\ 10 & 1 \end{vmatrix} + (1)\begin{vmatrix} 0 & 4 \\ 10 & 2 \end{vmatrix}$$

$$= 5(4 - 14) - 2(0 - 70) + (0 - 40) = 50 \quad (6c)$$

$$\begin{vmatrix} 5 & 2 & -6 \\ 0 & 4 & -1 \\ 10 & 2 & 3 \end{vmatrix}$$

$$= (5)\begin{vmatrix} 4 & -1 \\ 2 & 3 \end{vmatrix} - (2)\begin{vmatrix} 0 & -1 \\ 10 & 3 \end{vmatrix} + (-6)\begin{vmatrix} 0 & 4 \\ 10 & 2 \end{vmatrix}$$

$$= 5(12 + 2) - 2(0 + 10) - 6(0 - 40) = 290 \quad (6d)$$

Hence

$$|A| = (-2)(-90) + (1)(50) - (3)(290) = -640$$

It can be proved from Eq. (5) that the determinant $|A|$ is the sum of all terms of the form $(-1)^j a_{1j_1}, a_{2j_2}, \ldots, a_{nj_n}$ for all possible orderings $j_1, j_2, \ldots, j_n$ of the second subscripts 1, 2, ..., n, and the number j is the number of interchanges of two digits required to carry the ordering $j_1, j_2, \ldots, j_n$ into the natural ordering 1, 2, ..., n. Thus $|A|$ is the sum of all products of n elements, one from each row and one from each column, with a certain sign affixed to each product. Since there are $n! = 1 \cdot 2 \cdot 3 \ldots (n-1) \cdot n$ orderings of the numbers 1, 2, ..., n, there are $n!$ terms in the sum.

In the definition of a determinant in Eq. (5), A_{1j}, $j = 1, 2, \ldots, n$, which is a determinant of order $n - 1$, is called the complementary minor of the element a_{1j} of the first row. In general, the complementary minor of any element a_{ij} is the determinant of order $n - 1$, A_{ij}, which is formed from A by deleting the ith row and jth column. The cofactor of the element a_{ij} is $(-1)^{i+j}A_{ij}$. Thus in Eq. (5), the value of $|A|$ is given as the sum of the products of each element of the first row and its cofactor. It can be shown that for any row i, $|A| = a_{i1}(-1)^{i+1} A_{i1} + a_{i2}(-1)^{i+2} A_{i2} + \cdots + a_{in}(-1)^{i+n} A_{in}$. This expansion of $|A|$ is called the expansion of $|A|$ according to the elements of the ith row. Thus in Eq. (6),

$$
\begin{aligned}
|A| = (0)\,(-1)^{3+1}&\begin{vmatrix} 0 & 1 & 3 \\ 2 & -6 & 1 \\ 2 & 3 & 1 \end{vmatrix} \\
&+ (4)(-1)^{3+2}\begin{vmatrix} -2 & 1 & 3 \\ 5 & -6 & 1 \\ 10 & 3 & 1 \end{vmatrix} \\
&+ (-1)(-1)^{3+3}\begin{vmatrix} -2 & 0 & 3 \\ 5 & 2 & 1 \\ 10 & 2 & 1 \end{vmatrix} \\
&+ (7)(-1)^{3+4}\begin{vmatrix} -2 & 0 & 1 \\ 5 & 2 & -6 \\ 10 & 2 & 3 \end{vmatrix} = -640
\end{aligned}
$$

is the expansion of $|A|$ according to the third row.

The transpose of the square matrix A is the matrix, denoted by A', which has as its ith row the ith column of A and as its jth column the jth row of A for all i,$-j$. It can be proved that $|A'| = |A|$. It follows from this property that any theorem concerning the rows of a determinant implies a corresponding result concerning the columns, and vice versa. For example, a determinant can be evaluated by an expansion according to the jth column, by which $|A|$ is the sum of the products of each element of the jth column and its cofactor.

The following further properties of determinants follow from Eq. (5):

(i) If B is a matrix obtained from A by interchanging two rows (columns) of A, then $|B| = -|A|$.

(ii) If two rows (columns) of A are identical, then $|A| = 0$.

(iii) If B is a matrix obtained from A by multiplying every element of one row (column) of A by the number m, then $|B| = m|A|$.

(iv) If B is a matrix obtained from A by adding to each element of a row (column) of A a constant multiple of the corresponding element of another row (column) of A, then $|B| = |A|$.

(v) If B is a matrix identical with A except possibly for the ith row $b_{i1}, b_{i2}, \ldots, b_{in}$, and if C is a matrix identical with A except that the ith row of C is $a_{i1} + b_{i1}, a_{i2} + b_{i2}, \ldots, a_{in} + b_{in}$, then $|A| + |B| = |C|$. (There is the corresponding property for columns.)

These properties are used to shorten computation in finding the value of a determinant. In Eqs. (6),

$$
\begin{aligned}
|A| &= \begin{vmatrix} -2 & 0 & 1 & 3 \\ 5 & 2 & -6 & 1 \\ 0 & 4 & -1 & 7 \\ 10 & 2 & 3 & 1 \end{vmatrix} \\
&= 2\begin{vmatrix} -2 & 0 & 1 & 3 \\ 5 & 1 & -6 & 1 \\ 0 & 2 & -1 & 7 \\ 10 & 1 & 3 & 1 \end{vmatrix} \text{ by (iii) applied to the second column} \\
&= 2\begin{vmatrix} -2 & 0 & 1 & 3 \\ 5 & 1 & -6 & 1 \\ -10 & 0 & 11 & 5 \\ 10 & 1 & 3 & 1 \end{vmatrix} \text{ by (iv) multiplying the second row by } -2 \text{ and adding to the third row} \\
&= 2\begin{vmatrix} -2 & 0 & 1 & 3 \\ 5 & 1 & -6 & 1 \\ -10 & 0 & 11 & 5 \\ 5 & 0 & 9 & 0 \end{vmatrix} \text{ by (iv) multiplying the second row by } -1 \text{ and adding to the fourth row} \\
&= 2(1)(-1)^{2+2}\begin{vmatrix} -2 & 1 & 3 \\ -10 & 11 & 5 \\ 5 & 9 & 0 \end{vmatrix} \text{ by the expansion according to the elements of the second column} \\
&= 2\begin{vmatrix} 0 & 1 & 0 \\ 12 & 11 & -28 \\ 23 & 9 & 27 \end{vmatrix} = 2(1)(-1)^{1+2}\begin{vmatrix} 12 & -28 \\ 23 & -27 \end{vmatrix} \\
&= -8\begin{vmatrix} 3 & -7 \\ 23 & -27 \end{vmatrix} \\
&= -8(-81 + 161) \\
&= -640 \text{ by (iv), (iii), and Eq. (5)}
\end{aligned}
$$

If A and B are square matrices of order n, then the product AB is a square matrix of order n, and the element in the ith row and jth column, for all i, j, is obtained by multiplying the n elements in the ith row of A into the n elements in the j column of B, term by term, and adding these products. A most useful property of determinants is the fact that the determinant of the product AB is equal to the product of the determinant of A and the determinant of B; that is, $|AB| = |A|\,|B|$.

In the matrix A, select any r rows and r columns ($r \leq n$). The elements common to these rows and columns form an r by r matrix M, and the determinant $|M|$ is called an r-rowed minor of A. The determinant of the $n - r$ by $n - r$ matrix of elements common to the remaining $n - r$ rows and columns of A is called the complementary minor of $|M|$. When

$r = 1$ and row i and column j are selected, then $|M| = a_{ij}$ and the complementary minor is A_{ij}, which was defined earlier. The Laplace expansion of $|A|$ gives the value of $|A|$ as a sum, with certain signs, of all possible minors formed from a fixed set of r rows multiplied by their complementary minors. If $i_1, i_2, \ldots, i_r$ are the fixed rows, then the term in the sum, which is the minor formed from the columns $j_1, j_2, \ldots, j_r$ multiplied by its complementary minor, has the sign $(-1)^{I+J}$, where $I = i_1 + i_2 + \cdots + i_r$ and $J = j_1 + j_2 + \cdots + j_r$. In Eqs. (6), choosing the second and third rows as the fixed rows gives

$$\begin{aligned}
|A| = {} & (-1)^{2+3+1+2} \begin{vmatrix} 5 & 2 \\ 0 & 4 \end{vmatrix} \begin{vmatrix} 1 & 3 \\ 3 & 1 \end{vmatrix} \\
& + (-1)^{2+3+1+3} \begin{vmatrix} 5 & -6 \\ 0 & -1 \end{vmatrix} \begin{vmatrix} 0 & 3 \\ 2 & 1 \end{vmatrix} \\
& + (-1)^{2+3+1+4} \begin{vmatrix} 5 & 1 \\ 0 & 7 \end{vmatrix} \begin{vmatrix} 0 & 1 \\ 2 & 3 \end{vmatrix} \\
& + (-1)^{2+3+2+3} \begin{vmatrix} 2 & -6 \\ 4 & -1 \end{vmatrix} \begin{vmatrix} -2 & 3 \\ 10 & 1 \end{vmatrix} \\
& + (-1)^{2+3+2+4} \begin{vmatrix} 2 & 1 \\ 4 & 7 \end{vmatrix} \begin{vmatrix} -2 & 1 \\ 10 & 3 \end{vmatrix} \\
& + (-1)^{2+3+3+4} \begin{vmatrix} -6 & 1 \\ -1 & 7 \end{vmatrix} \begin{vmatrix} -2 & 0 \\ 10 & 2 \end{vmatrix} \\
= {} & (20)(-8) - (-5)(-6) + (35)(-2) \\
& + (22)(-32) - (10)(-16) + (-41)(-4) \\
= {} & -640
\end{aligned}$$

The Laplace expansion is particularly useful in evaluating the determinants of matrices which have blocks of zeros, for example, using the first three rows as the fixed rows,

$$\begin{vmatrix} -5 & 4 & -1 & 0 & 0 & 0 \\ 2 & 6 & 3 & 0 & 0 & 0 \\ 0 & 1 & -1 & 0 & 0 & 0 \\ 0 & 0 & 0 & 4 & -1 & 0 \\ 0 & 0 & 0 & 2 & 7 & 1 \\ 0 & 0 & 0 & 0 & 5 & 3 \end{vmatrix}$$

$$= \begin{vmatrix} -5 & 4 & -1 \\ 2 & 6 & 3 \\ 0 & 1 & -1 \end{vmatrix} \begin{vmatrix} 4 & -1 & 0 \\ 2 & 7 & 1 \\ 0 & 5 & 3 \end{vmatrix}$$

$$= (51)(70) = 3570$$

Every minor formed from the first three rows is zero by property (ii), except the one formed from the first three columns. *See* EQUATIONS, THEORY OF.

Ross A. Beaumont

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Determinism

The principle that nature follows exact laws, so that what will happen in the future is a necessary consequence of the state of the world at any given moment in the past. This view, if fully adopted, implies that events which seem to occur by chance would be fully understood if more was known about them, and that apparently free thoughts and choices are explainable and in principle predictable in terms of neuroscience. In a looser sense, determinism refers to claims that mental freedom is much more restricted than is generally supposed. An immense philosophical literature has arisen in response to this challenge to the sense of humanity, most of it defending the freedom of a person's choices and the necessity that the individual be held responsible for them.

The question of determinism in physical science cannot be considered apart from the philosophical problem; this gives it added importance and forces its consideration in a very critical spirit.

Prequantum ideas. The idea that the world is composed of atoms moving under the influence of certain forces according to certain laws can be traced back to the Greek philosopher Leucippus, who wrote in the fifth century B.C., "Nothing happens at random but everything for a reason and by necessity." In subsequent centuries this idea of necessity was never entirely lost. It was prominent in the work of the seventeenth-century thinker René Descartes, and became widely known through his influence. Isaac Newton carried out a large part of the cartesian scientific program. His theory explained so many natural processes that it began to appear that the universe since the time of Creation might actually have run its course in a deterministic fashion like a machine, without divine intervention. Newton sought to avoid this conclusion by suggesting that gravitation may be an act of God, and that at any rate He must continually intervene to keep the starry world from collapsing under the gravitational force. The negative implications for free will were widely considered during the seventeenth century, and Descartes was among those who made suggestions for avoiding them.

A century after Newton, Pierre Simon de Laplace argued that an Omniscient Calculator, provided with exact knowledge of the state of the universe at present, would be able to predict the entire future. *See* CLASSICAL MECHANICS; NEWTON'S LAWS OF MOTION.

Quantum physics. In the early 1900s a reaction began against an exclusively mechanical explanation of nature. It took the form of a largely unjustified condemnation of atomic theory on the grounds that the proper concern of science is with things which can be observed and which do not have to be imagined. The great successes of atomic theory soon put an end to this skepticism, but enough of it remained to make easier the later renunciation of determinism on the atomic scale.

The quantum mechanics of the 1920s opened the way to a reassessment of determinism in nature when it introduced the paradox of particles which are, at

the same time, waves. A wave of length λ was supposed to accompany, or describe, a particle of momentum $p = h/\lambda$, where h is Planck's constant. But a wave with only the single Fourier component λ extends over all space from $-\infty$ to $+\infty$ and says nothing about the location of the particle. The solution to this problem was given by Fourier's superposition of waves and Max Born's probabilistic interpretation of quantum mechanics: express the probability of finding a free particle by a (complex) wave packet (**Fig. 1**). This can be represented by a superposition of waves with wavelengths in the neighborhood of λ, and according to Born the particle has appreciable probability of being found only where the wave function ψ (or more precisely, $|\psi|^2$) is large; that is, within a region of size roughly Δx on each side of x_0 in Fig. 1. It is convenient to introduce the quantity $k = 2\pi/\lambda$, in terms of which the relation $p = h/\lambda$ becomes $p = \hbar k$, where $\hbar = h/2\pi$. A theorem of Fourier analysis shows that Δk, the range of k's present in the wave packet, is related to Δx by relation (1). Each side of relation (1) can be multiplied by $\hbar$ to get relation (2). This states that the more

$$\Delta x\ \Delta k \geq 1 \tag{1}$$

$$\Delta x\ \Delta p \geq \hbar \tag{2}$$

precisely the specification of a particle's position x is attempted by means of a localized wave function (the only means available in quantum mechanics), the less precisely can its momentum be specified. *See* FOURIER SERIES AND TRANSFORMS.

Relation (2), a trivial consequence of Fourier's theory, was written down by Werner Heisenberg in 1927. It immediately raised the question of whether the inability of the theory to make more precise specifications was a blemish on the theory or whether it reflected some previously unsuspected property of nature. Heisenberg broke the three-century-old tradition of Descartes by choosing the second alternative. His principle of uncertainty, or indeterminacy, states that it is impossible to measure, and therefore to know, x and p any more accurately than is allowed by relation (2), and there are also other pairs of dynamical variables similarly related.

Heisenberg and Niels Bohr worked out an example to illustrate the uncertainty principle (**Fig. 2**). The least possible amount of light is a single quantum, a photon. Imagine an idealized optical arrangement in which this is enough to locate the particle P. The

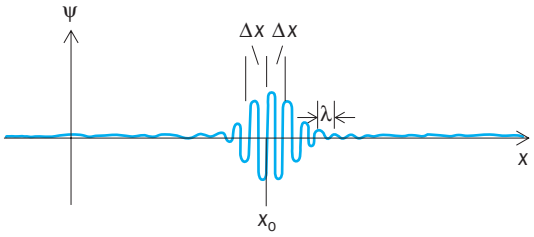

Fig. 1. A wave packet—of wavelength approximately λ, and produced perhaps by opening and closing a shutter—traveling through space.

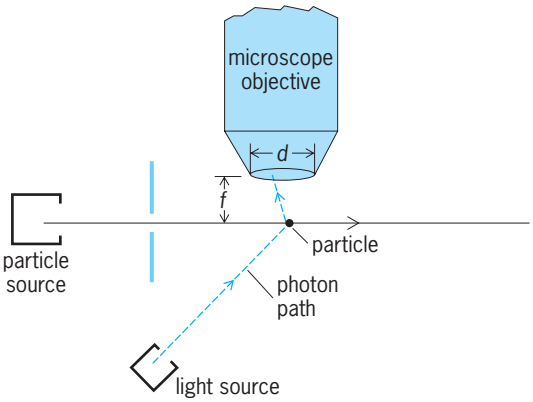

Fig. 2. Example of the uncertainty principle.

quantum emerges from the light source and bounces off P and into the lens of the microscope. In doing so, it changes the particle's momentum. If the quantum could be located after it hit P, it would be possible to allow for this recoil; but it is only known that the quantum entered the lens. The uncertainty in the recoil produces an uncertainty Δp in the particle's subsequent momentum given by relation (3), where

$$\Delta p \geq p\frac{d}{f} = \frac{hd}{\lambda f} \tag{3}$$

p is the momentum of the photon, d is the diameter of the lens, f is the distance from the particle to the lens, and λ is the wavelength of the light emitted by the light source. But by the principles of optics a microscope cannot locate a particle perfectly; the finite wavelength of the light gives a fuzziness in the image which leads to an uncertainty in the particle's measured position given by relation (4).

$$\Delta x \geq \frac{\lambda f}{2\pi d} \tag{4}$$

Multiplying these relations together gives relation (2), independently of any choices made in designing and using the microscope. Many physicists have tried to circumvent relation (2) by inventing clever devices, but it has always turned out that there is a flaw in the reasoning. *See* UNCERTAINTY PRINCIPLE.

There remains the question as to whether Heisenberg's principle is merely an unfortunate limitation on an experimenter's ability to know or whether it goes deeper. The general opinion of physicists is that of Bohr: the principle expresses a limitation of the precision with which concepts such as position and momentum can be applied at all. Lacking these concepts, Laplace's Omniscient Calculator cannot predict the future. (Nevertheless, the laws of quantum mechanics are supposed to be exact. They refer to probabilities, however, and so would not help the Calculator.)

The fall of determinism produced a profound effect among scientists and philosophers. Some welcomed it; others denied it. Einstein made the much quoted remark, "I do not believe that God plays dice," but in general, scientific opinion accepts the new situation. Whether the human mind is freed

from the bonds of determinism by Heisenberg's principle is not known. Present understanding of mental functions is far too primitive to make any pronouncements on this question. *See* NONRELATIVISTIC QUANTUM THEORY; QUANTUM MECHANICS.

Newtonian physics: modern ideas. Perceptive mathematicians have warned that determinism is not as obvious a consequence of newtonian physics as it might appear. Henri Poincaré and Emile Borel, and, since a famous address by Nikolai Kolmogorov, a distinguished school of Russian mathematicians have proved results which must be stated mathematically, but whose general effect is that for the vast majority of dynamical systems any error in the initial conditions, however small, will be amplified, in general exponentially, and so quickly that the predicted result will soon bear no relation to reality. Thus unless it is assumed that initial conditions are known, and computation is performed, with perfect accuracy, the Omniscient Calculator will wind up getting everything wrong.

An example of this behavior (although not exponential) involves a ball that is dropped onto a knife edge. If the ball is dropped from a point Q a little to the left of the edge, it will end up on the left side of the knife. If Q is slowly shifted toward the right, a location will be reached at which the ball begins to fall to the right. Whether or not there exists an intermediate place Q_0 from which the ball will bounce up and down on the knife edge and finally come to rest there forever, it is apparent that very small departures on either side of the real or hypothetical Q_0 will cause the ball to end up in very different places. One result of the mathematical work is to show that similar instabilities occur in most dynamical systems even if there are no knife edges. *See* CATASTROPHE THEORY.

Very few people now think that all events in the natural world are exactly determined. Experiments suggest that some human and animal behavior can reliably be predicted and controlled, but nobody knows the limits within which this can be done. David Park

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Deuterium

The isotope of the element hydrogen with atomic weight 2.0144 and symbols ^{2}H or D. Considerations of nuclear stability and a discrepancy between the chemical and physical atomic weights of hydrogen led to the prediction of a stable isotope of hydrogen of mass 2. A successful search for this isotope, deuterium, was made by H. C. Urey, F. G. Brickwedde, and G. M. Murphy in 1931. The terrestrial natural abundance of deuterium is 1 in 6700 parts of ordinary hydrogen (protium), of atomic weight 1.0078. Small variations in natural sources are found as a result of fractionation by geological processes. Industrial hydrogen, particularly that generated by electrolysis of water, may contain significantly less deuterium.

Deuterium is used mainly in the form of heavy water. In the uncombined state it finds uses as a research tool. Liquid deuterium is used in bubble chambers to study the reactions of elementary particles with the deuterium nucleus, the deuteron. Deuterons are frequently accelerated in cyclotrons to study their reactions with other nuclei and also to produce radioactive nuclides. Deuterium gas is used in the direct synthesis of organic compounds for tracer studies. Illustrative of such procedures are exchange reactions with hydrogen-containing substances in the presence of hydrogenation catalysts. If controlled thermonuclear fusion can be achieved, deuterium gas would become an exceedingly important source of power. *See* NUCLEAR FUSION; NUCLEAR REACTION.

Deuterium, D_2, is a gas at room temperature. It is prepared from heavy water, D_2O, either by electrolysis or by reaction of D_2O with metals such as zinc, iron, calcium, and uranium. It is also prepared directly by the fractional distillation of liquid hydrogen. In this process it is necessary to catalyze the disproportionation (unsymmetrical dissociation and recombination) of HD into H_2 and D_2 in order to obtain D_2.

Ordinary deuterium is a molecular mixture of two-thirds *ortho*- and one-third *para*-deuterium. At 20 K the equilibrium composition is 97.8% *o*-deuterium. Deuterium molecules obey the Bose-Einstein statistics, and the ortho species have even rotational quantum numbers, whereas the para species have odd rotational quantum numbers. The analysis of the ortho-para composition of deuterium is most conveniently made by measurement of the thermal conductivity of the gas at 77 K. The physical and chemical properties of *o*- and *p*-deuterium are very similar. In most of its physical and chemical properties, deuterium resembles protium. In most cases, deuterium is slightly less reactive than protium. An intercomparison of some of the physical properties of D_2 with those of H_2 is given in the **table**.

Deuterium gas is usually slightly contaminated with HD. The analysis of the gas for protium is most conveniently carried out by a mass spectrometer. Alternative methods of analysis are by thermal conductivity and the rate of effusion of the gas through an orifice. The gas can be converted to water, and the following properties have been used as a basis for analysis: infrared spectrum, density, index of refraction, and nuclear magnetic resonance.

The chemical reactivity of deuterium is less than that of hydrogen because of its lower zero-point energy and smaller collision frequency. At 1000 K deuterium is 32% less dissociated than protium. At room temperature deuterium atoms are electrolyzed out of water in the form of hydrogen gas, at one-eighth the rate of protium atoms. *See* DEUTERON; HEAVY WATER; HYDROGEN; TRITIUM. Jacob Bigeleisen

Selected values of some physical properties of deuterium

Property	*n*-H_2	*n*-D_2	97.8% *o*-D_2
Triple point	13.96 K	18.72 K	18.63 K
Normal boiling point	20.4 K	23.6 K	23.5 K
Critical temperature	33.24 K	38.35 K	
Critical pressure	12.8 atm	16.4 atm	
Heat of fusion	28.0 cal/mole	47.0 cal/mole	
Heat of vaporization (at normal boiling point)	216 cal/mole	293 cal/mole	
Molar volumes of liquid at 20 K	28.3 ml	23.5 ml	

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Deuteromycota

A heterogeneous group of anamorphic (asexual or imperfect) fungi in which sporulation may occur on separate hyphae or composite fruit bodies (conidiomata). These may or may not be differentiated into distinct conidiophores and/or conidiogenous cells (conidium-bearing hyphae or cells) which occur separately or are situated over the surface of or inside a conidioma. Some lack conidiomata and do not produce conidia but form other types of propagules. Conidium-, propagule-, or conidioma-bearing septate hyphae, or individual cells, constitute the vegetative thallus. The diagnostic feature of the group is the lack of a teleomorphic (sexual or perfect) state. There are more than 2500 accepted genera containing 21,000 species.

Even though teleomorphs are absent or have not been correlated in many members of the group, it is possible to determine the teleomorphic affinities of some deuteromycetes by the presence or absence of clamp connections, differences in septal structure, and varying proportions and nature of cell wall polysaccharides. More recently the use of molecular techniques in documenting genomes has facilitated the recognition of monophyletic groups. Most have affinities with Ascomycotina and a smaller number with Basidiomycotina. The anamorphs of Zygomycotina, Erysiphales, and Urediniomycetes are readily recognizable for various reasons and are assignable to these teleomorphic groups. It has been postulated that the Deuteromycotina are a taxon of temporary nature that would gradually disappear as the missing teleomorphs are discovered and correlated. However, some anamorphic fungi have with time become isolated geographically, genetically, spatially, and nutritionally from ancestral teleomorphs. The divergent evolution of such anamorphs and the ancestral teleomorphs has resulted in a permanent residue of organisms distinct from their teleomorphic precursors but which may be correlated with them by using molecular characters. *See* ASCOMYCOTA; BASIDIOMYCOTA; ZYGOMYCOTA.

Reproduction. The Deuteromycotina are also artificial inasmuch as the group consists of fungi or states of fungi reproducing vegetatively by mitosis, not meiosis. Some organisms, such as *Aspergillus, Penicillium*, and *Fusarium* have a parasexual process in which reciprocal exchange of chromatid segments between homologous chromosomes during mitotic propagation of the diploid nucleus occurs. Instead of the regular sequence of events which obtain in the meiotic cycle, the various processes may all be occurring at one time in one mycelium. For deuteromycetes, parasexuality brings the advantage of genetic recombination. *See* MITOSIS; RECOMBINATION (GENETICS).

Classification. Traditionally, Deuteromycotina are separated into two classes: Hyphomycetes, which are mycelial forms bearing conidia on separate hyphae or aggregations of hyphae, always external to the supporting substratum not inside discrete conidiomata; and Coelomycetes, which are forms that produce conidia in pycnidial, pycnothyrial, acervular, cupulate, or stromatic conidiomata. A third class is sometimes recognized, the Agonomycetes (Mycelia Sterilia); these fail to produce conidia and thus lack conidiomata, although they do form somatic survival propagules. Currently, no ranks are distinguished within the Deuteromycotina, and the subdivisional name itself is no longer in use, the preferred term being Mitotic Fungi or Fungi Anamorphici. The current trend which will obtain for the forseeable future is for deuteromycetes to be incorporated into teleomorph classifications by using traditional morphological approaches combined with molecular systematics. *See* AGONOMYCETES; COELOMYCETES; HYPHOMYCETES.

Asporogenous (imperfect) yeasts have sometimes been separated into the Blastomycetes, which are divided into two orders—the Cryptococcales, which have both ascomycete and basidiomycete affinity and reproduce by budding only, and the Sporobolomycetales, which have a basidiomycete affinity and reproduce by both budding and ballistoconidia. *See* BLASTOMYCETES.

Commercial utilization. Deuteromycetes are ubiquitous and occupy most conceivable ecological niches, whether they are aquatic, terrestrial, or artificial. Deuteromycetes have been used extensively in manufacturing processes of different kinds. Fermentation by yeasts and enzyme production by filamentous fungi such as *Aspergillus* is the basis for the alcoholic beverage industry and many Asian

specialty food products. Various *Penicillium* species are used in the production of blue cheeses. Many deuteromycetes produce secondary metabolites and are used in antibiotic and organic acid production processes. Fungal protein is manufactured by using species of *Fusarium, Trichoderma*, and *Verticillium;* other fungal biomass, from yeasts such as *Candida* and filamentous fungi such as *Paecilomyces* and *Sporotrichum*, is produced by utilization of paper mill liquid wastes. Plant wastes or residues, such as cereal straw, can be treated by cellulytic deuteromycetes, such as *Trichoderma viride*, resulting in liberated glucose. The product is then extracted and dried, and the unhydrolyzed residues are used either as adjuncts to animal feed, for fuel, or for soil amendments in agriculture and horticulture. *Penicillium, Aspergillus*, and *Saccharomyces* are sometimes used for the extraction and recovery of metals from natural sources and industrial effluents. In natural habitats the deuteromycetes are primarily responsible for the conditioning or mineralization of plant litter. Increasingly important is the use of deuteromycetes in biological control of insects, nematodes, pathogenic fungi, and weeds. *See* DISTILLED SPIRITS; FOOD MANUFACTURING.

Medical mycology. All fungal infections of humans are opportunistic, and mycoses are almost always the result of abnormal host immunity. Even before the association of the HIV virus with acquired immune deficiency syndrome (AIDS), criteria used in defining the syndrome relied heavily on the existence of certain opportunistic infections in the absence of any of the established predisposing conditions. A significant number of these deep-seated infections are caused by deuteromycetes such as *Aspergillus, Candida, Cryptococcus*, and *Histoplasma*. Superficial mycoses due to dermatophytes such as *Trichophyton* are commoner in patients with AIDS. Deuteromycetes also cause numerous diseases of plants; these are often of considerable economic importance, as are the losses caused in stored natural and manufactured products and processed food. In addition, mycotoxins may be produced directly by growth of *Fusarium, Penicillium*, and *Aspergillus* and many others on animal feed or human food. Such mycotoxins may pass through the food chain to animal products such as milk or meat, and illnesses arising from such sources are termed secondary mycotoxicoses. *See* ACQUIRED IMMUNE DEFICIENCY SYNDROME (AIDS); MEDICAL MYCOLOGY; MYCOTOXIN; OPPORTUNISTIC INFECTIONS; PLANT PATHOLOGY.

B. C. Sutton

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Deuteron

The nucleus of the atom of heavy hydrogen, H^2 (deuterium). The deuteron *d* is composed of a proton and a neutron. As the simplest multinucleon nucleus, the deuteron has been the subject of extensive study. Its binding energy is 2.227 MeV; that is, this is the amount of energy which must be added to a deuteron for it to dissociate into a proton and a neutron. The accurate determination of this dissociation energy provides the means of calculating the mass of the neutron, the mass of the deuteron (2.014187 amu) and proton being known from other experiments.

The intrinsic angular momenta, or spins, of the proton and neutron combine to produce a deuteron spin of unity; hence, the deuteron obeys the type of quantum statistics that is known as Bose-Einstein statistics. The deuteron possesses a magnetic moment (0.857407 nuclear magneton) and an electric quadrupole moment (2.738×10^{-27} cm^2).

Deuterons are much used as projectiles in nuclear bombardment experiments, especially to produce (d,p), (d,n), and (d,α) reactions. In the first two reactions, because of the low binding energy of the deuteron, the neutron *n* or proton *p* is stripped from it and captured by the target nucleus. Meanwhile, the other half of the deuteron (that is, the proton or neutron) carries away the excess energy. The H^1/H^2 abundance ratio in nature is 6700. *See* DEUTERIUM; NUCLEAR REACTION.

Henry E. Duckworth

Deuterostomia

A major division of the animal kingdom comprising the phyla Hemichordata, Echinodermata, and Chordata. Hemichordata consists of acorn worms (Enteropneusta) and a group of small-bodied, tentaculate forms (Pterobranchia). Echinodermata contains the starfish, sea urchins, crinoids (sea lilies), and their allies. The phylum Chordata includes the Urochordata (sea squirts and larvaceans), Cephalochordata (*Branchiostoma*, the sand lancelet, formerly *Amphioxus*), and vertebrates (see **illustration**). At one time, up to 10 phyla were included in Deuterostomia on the basis of shared morphological and developmental features; however, comparison of deoxyribonucleic acid (DNA) sequences indicates that most of those phyla are more closely allied to other groups. *See* ANIMAL KINGDOM.

General characteristics. A number of morphological and developmental features are widely shared among deuterostomes; for example, the adult mouth always forms at some distance from the blastopore (the opening into the embryonic gut), a condition termed deuterostomy, which gave the group its name. In many (but not all) deuterostomes, early embryos show radial cleavage and form coelomic body cavities (which house the internal organs) from outpockets of the embryonic gut. However, none of those features is exclusive to the group, which means that they were either present before deuterostomes evolved (and thus were inherited by other

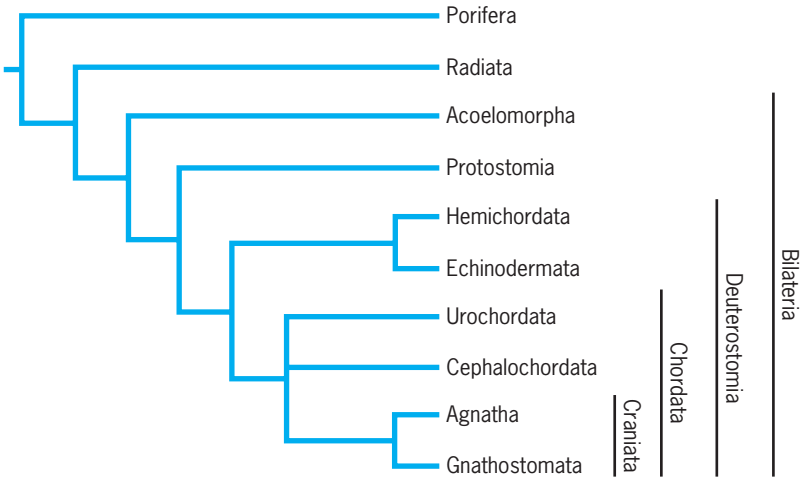
groups as well) or have evolved in other groups independently. One feature that is exclusive to deuterostomes, though not present universally among them, is pharyngotremy, characterized by a perforation of the pharynx that permits an outflow of water that has been ingested through the mouth, usually bathing gills; it is likely a primitive character (heritable trait) of this crown group (that is, the last common ancestor of all living deuterostomes, plus all of its descendants).

Hemichordata. Molecular phylogenetic studies indicate that hemichordates and echinoderms are more closely related to each other than either is to chordates. Among hemichordates, the benthic, vermiform enteropneust body plans are the least derived (advanced): their coeloms are divided into three successive anteroposterior regions (the proboscis, collar, and trunk), and the nervous system consists of a subepidermal plexus that evidently provides most of the integration and coordination, with supplementary dorsal and ventral nerve cords. *See* HEMICHORDATA.

Echinodermata. The echinoderms are highly derived, and while three anteroposterior coelomic regions are also present during early development, they become modified, partly conjoined, and partly lost as the developing body plan produces the adult pentameral symmetry found in starfish and most other living classes. The echinoderm nervous system is also a subepidermal plexus. Echinoderm bodies are partly supported and protected by a carbonate endoskeleton of spicules or, commonly, fused plates. Living echinoderm groups do not show pharyngotremy, but extinct primitive groups, which lacked pentameral symmetry, may have displayed this characteristic. *See* ECHINODERMATA.

Chordata. The chordates evidently share a common vermiform ancestor with enteropneusts, but chordate ancestors evolved adaptations for swimming that led to an innovative body plan. Chordates evolved a springy dorsal structure, the notochord, which antagonized lateral swimming motions that were produced by anteroposteriorly arranged muscle blocks (myotomes), as well as a lateral flattening of the body, unusual in organisms living in benthic habitats but common among active swimmers. Coordination of functions was consolidated in the dorsal nerve cord, presumably positioned in association with the notochord so as to innervate the series of muscle blocks, forming a segmental system.

Acraniata and Craniata. Several important clades (groups of related organisms descended from a common ancestor) radiated from early chordates. One is Acraniata, containing the cephalochordates, particle feeders that lack axial skeletons. Another major clade is Craniata, which include agnathan vertebrate groups, jawless forms with cartilaginous skeletons represented today by hagfish and lampreys. The earliest of known Cambrian vertebrates, from approximately 520 million years ago (mya), was evidently an agnathan, as probably were the euconodonts, extinct eel-like forms having cranial cartilage and furnished with a feeding apparatus that included the first signs of bone, dating from nearly 510 mya or slightly later. Making up the bulk of Craniata today are the gnathostomes, jawed vertebrates that are often divided into seven Linnean classes that include the well-known groups of living vertebrates.



Position of Deuterostomia and some of its major divisions within the animal kingdom.

Urochordata. Another major deuterostome clade, with a considerable diversity of living marine species (about 1250), is Urochordata. It is often regarded as a phylum in its own right, but more recently has been considered a subphylum of Chordata, implying that it was part of an early chordate radiation, but as a separate branch from Acraniata and Craniata. Living members of Urochordata appear to be so highly modified and probably simplified from their ancestral condition that the ancestry of urochordates and their place in the evolution of other chordate clades has been obscured. The most obvious deuterostome characteristic of urochordates is a large, highly perforate pharynx, used as a feeding basket for filtering food items from seawater. The relation of urochordates to the chordates is indicated chiefly by the presence of a notochord in the larvae. *See* CHORDATA; VERTEBRATA.

Evolution. Molecular clock evidence (an estimate of the time of divergence of a lineage based on the rate of genetic change) suggests that the last common ancestor of deuterostomes lived over 560 mya, with the trace fossil record suggesting that this ancestor was a small bilaterian worm. The earliest undisputed deuterostome body fossils, probably dating back 525 mya, are probably extinct classes of echinoderms, easily preserved because of their durable skeletons. Spectacular fossil deposits from China, probably dating nearly 520 mya, have yielded abundant remains of soft-bodied animals, and among them are urochordates and even vertebrates, which are represented by cartilaginous forms that are probably allied to living agnathans.

Molecular studies of deuterostome genes and genomes has produced evidence of important stages in their evolution. For example, during development the genes of the *Hox* cluster help mediate the pattern of anteroposterior differentiation in all bilaterians

investigated. *Hox* genes are transcriptional regulators with homeobox motifs (DNA sequences that regulate somatic development). The last common ancestor of crown deuterostomes had at least 7 *Hox* genes. Echinoderms have 10 *Hox* genes, while cephalochordates have 14, possibly the primitive chordate number. Vertebrates have multiple clusters that have arisen by cluster duplications, 4 for mammals and 7 for some fish, but each cluster has lost many genes so that the total number of Hox genes in mammals is 39 rather than 60—each cluster averages about 10 genes. Thus in general, in the deuterostome genome a history of increasing body plan complexity is reflected by increasing numbers of *Hox* genes. However, the known genome of urochordates includes only 9 *Hox* genes; either it is primitive with respect to cephalochordates or has been simplified during evolution. *See* ANIMAL EVOLUTION; DEVELOPMENTAL GENETICS; HOMEOSIS. James W. Valentine

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Developmental biology

A large field of investigation that includes the study of all changes associated with an organism as it progresses through the life cycle. The life cycles of all multicellular organisms exhibit many similarities. That is, as an organism progresses from one generation to the next there is a series of common processes: for example, gametogenesis, fertilization, embryogenesis, cell differentiation, tissue differentiation, organogenesis, maturation, growth, reproduction, senescence, and death. *See* ANIMAL GROWTH; CELL CYCLE; CELL SENESCENCE AND DEATH; EMBRYOGENESIS; FERTILIZATION; GAMETOGENESIS; PLANT GROWTH.

History

In the early nineteenth century, when professional science was emerging, the term biology was first used to include botany and zoology with agricultural and medical applications. Embryology, the first subdiscipline devoted to development, was a branch of anatomy, which at the time was concerned with organs and tissues, with pathology and teratology as sidelines. Embryology concerned itself with the observation and delineation of the stages by which an embryo was shaped into adult form: progressive changes in its size and in its organs, tissues, and cells. At the same time, the nucleate cell was being recognized as the basic unit in living plants and animals. The tiny and elusive mammalian ovum was discovered in this period.

In the second half of the century, biological discoveries abounded. Cell studies flourished, encouraged by improvements not only in optics but in other laboratory equipment and in chemicals, especially dyes. Marine organisms furnished ideal living material for the study of cellular stages in animal development. Laboratories for efficient work became organized. By the 1880s an experimental branch of embryology was an established fact: its concern was the mechanics of development.

By the end of the nineteenth century, the structural aspects of cell division (mitosis) had been observed in plant and animal tissues, including the accurate division of the chromosomes by the spindle apparatus. Fertilization of egg by sperm, with the union of the two nuclei, was recognized. Finally, meiosis, whereby the double (diploid) chromosome set was reduced to a single (haploid) condition, was elucidated. By the turn of the century, the species specificity of the chromosome sets was clear, and also the individuality of each chromosome. Almost from the time of its discovery, the chromatin apparatus of the nucleus was seen to have something to do with heredity; the rediscovery of Mendel's neglected experiments offered an obvious corroborative pattern. *See* CHROMOSOME; MEIOSIS; MITOSIS.

In the meantime the experimentalists were investigating the egg and the early stages of its development, not only in detailing cell lineages in the cleavage stages that followed insemination, but in manipulations such as hybrid fertilization, and in observing the developmental potentialities of parts of the egg or of separated blastomeres, and the direction of the cleavage pattern by pressure. External factors also were investigated: physical or chemical activation of the egg in the absence of sperm, and various developmental roles of calcium in seawater. In inland laboratories, microsurgery of vertebrate embryos was explored.

The twentieth century brought continuous expansion of experimental techniques. Microsurgery, particularly of the amphibian embryo, became a fine art. The centrifuge was used for rearranging cellular inclusions; micromanipulators were devised for surgery and injection under very high magnification; physiological apparatus was adapted for very small objects. Photomicrography was continually improved, as well as time-lapse cinematography, a great convenience in condensing a long life history for review in limited time. Optical study of living cells was greatly advanced. Dark-field and polarizing optics had long been known; interference and phase-contrast optics were developed, and electron microscopy was considered. Tissue culture was invented as a device for following cellular differentiation outside the body, and later was adapted for observation of organ and embryo development. Only much later were methods found to culture dissociated cells, and to induce cell-to-cell fusion and even hybridization of somatic cells. *See* MICROMANIPULATION; MICROSCOPE; TISSUE CULTURE.

Genetics as a science dates from about 1900. The close association between embryology, cell biology, and genetics was obvious, but functional amalgamation of these disciplines did not occur at once.

The pioneer geneticists were preoccupied with hybridization and chromosome study, and were intent on establishing a sound theory of the gene. By the 1920s, however, several distinct branches of genetics could be identified: breeding, usually associated with chromosome cytology; population genetics, associated with evolutionary, ecological, and sociopolitical orientation; and physiological genetics, which soon became developmental genetics as information accumulated relevant to the mechanism of expression of known genes. Lethal genes, as an extreme form of teratological factors, or "inborn errors of metabolism," are particularly useful for investigation of developmental processes. *See* CELL BIOLOGY; EMBRYOLOGY; GENETICS; METABOLIC DISORDERS.

In the 1930s a new trend emerged: microbial genetics. At first, eukaryote (nucleate) species with sexual cycles (for example, molds and yeasts) were shown to possess orthodox genetic systems controlling fundamental biochemical properties such as nutritional requirements and enzyme production. Eventually the prokaryote bacteria, whose possession of genes had been in question because no formed nucleus or chromosomes had been recognized, were shown to be capable of genetic exchange; they now furnish indispensable laboratory material. *See* BACTERIAL GENETICS; GENETIC ENGINEERING.

Biochemistry in this period, profiting from the almost revolutionary surge of atomic physics, was learning a great deal about macromolecules (proteins, nucleic acids, polysaccharides, lipids), to the study of which crystallographic methods were being applied. Fundamental metabolic reaction systems were beginning to be understood as serial or cyclic arrangements of substrates and enzymes. Radioactive isotopes were becoming available, to label reagents and eventually to locate reaction products within cells. Ingenious methods were devised for the separation of cellular constituents and identification of substances in very small samples. The search for the genetic material, narrowed to either protein or nucleic acid, was decided in favor of the latter. All these advances were well under way before World War II. *See* BIOCHEMISTRY.

By the end of that war, machinery had been devised for rapid analytic procedures that previously had been hopelessly tedious and time-consuming. The electron microsope was being developed and refined even so far as to resolve macromolecules, thus opening an entire new world to visual inspection. Microbiological and cell-culture methods provided experimental material supplementing large organisms, and served as excellent analytic tools. The two nucleic acids and their subunit composition were located: deoxyribonucleic acid (DNA), mainly confined to the nucleus of the cell, and ribonucleic acid (RNA), principally in the cytoplasm. The calculation of an acceptable model for DNA structure and subsequently the understanding of how the sequence of that structure could furnish a code, whereby RNA could designate protein structure (translation), constituted a spectacular step for molecular biology. The principal features of the DNA double-helix model were soon verified, and modified, by designed experiments on prokaryote material. Bacterial nucleic acid systems proved highly rewarding. Their code, designating the amino acid sequence in the primary protein assembly, was delineated, and their several functions ascertained. DNA, as predicted, acts as a template for its own replication, preceding cell division; RNA, transcribed from DNA, operates in at least three distinct forms to assemble amino acid sequences according to the transcribed code. Within a few years, protein synthesis was reconstructed in test tubes. Virus particles, with no synthesizing system of their own, were recognized as parasites consisting of a core of either DNA or RNA covered by a protein coat and capable of transferring their templates to the host cell. Bacterial as well as other DNA strands proved to break and rejoin in surprising combinations. *See* DEOXYRIBONUCLEIC ACID (DNA); ELECTRON MICROSCOPE; GENETIC CODE; MOLECULAR BIOLOGY; PROTEIN; RIBONUCLEIC ACID (RNA); VIRUS.

Dorothea Rudnick

Developmental View

As can be seen in the preceding discussion, developmental biology is a broad-ranging discipline which includes many areas of biological investigation. Thus analysis of all of the events associated with an organism as it progresses through its life cycle employs a multiplicity of approaches. In general the analysis has been reductionist in nature. Tremendous strides have been made in describing at the molecular level the developmental process of cell differentiation. However, the molecular control mechanisms which regulate cell differentiation are not known. Tissue and organ differentiation, as well as morphogenesis, are processes which have been described in detail for many situations, but little is known about the physical and chemical nature of the mechanisms involved. A complete understanding of the development of an organism will require an appreciation and comprehension of the changes which occur at all levels of organization as an organism traverses its life cycle.

The major unifying theme in biology is evolution. Not only has evolution led to the wide variety of organisms now present on Earth, but also evolution has modified the initial processes and patterns of development to the diversity of types currently encountered. This evolution of developmental parameters in multicellular organisms began as single-celled organisms became multicellular. The development of a multicellular organism entails a host of problems not faced by a single-celled organism. For example, cells in one part of the aggregate must coordinate their activities with cells in other parts, nutrients and oxygen must be provided to all cells, and water balance must be maintained. In the two major groups of multicellular organisms, plants and animals, there appear to be two differences which may account for their different solutions to the common problems associated with being multicellular. First, plants are autotrophic because they can fix carbon directly via

photosynthesis. Thus, they can remain in one spot and provide for their needs as long as light energy is available. Second, their cells are encased in a box, the cell wall, which prevents movement of cells in relation to each other and severely restricts overall movement of the organism and its parts. Thus, shape and form result primarily from cell division and expansion, and rarely from cell death. Animals, on the other hand, are usually motile and heterotrophic, obtaining food from their environment. In many instances animal cells are motile, and this makes possible additional mechanisms for attaining shape and form. With this viewpoint, one can understand many of the similarities and differences which have arisen in plants and animals as they evolved. *See* ANIMAL EVOLUTION; CELL DIVISION; CELL WALLS (PLANT); PHOTOSYNTHESIS; PLANT EVOLUTION.

Developmental biologists have focused on two central areas: the processes and associated mechanisms by which cells become different, that is, cell differentiation; and the processes and associated mechanisms by which patterns are created, that is, morphogenesis.

Totipotency. Current theories state that cells become different by expressing different genes. Thus, a liver cell is different from a muscle cell, not because it contains different genes or genetic information, but because it expresses different sets of genes. This explanation of cell differences is based upon the results of three types of experimental analysis. (1) Some plant cells are totipotent; that is, for tobacco, carrot, and a few other plant species, it has been demonstrated that a single cell (not a gamete) can divide and undergo morphogenesis to form a fertile plant. (2) Nuclei from some differentiated animal cells are totipotent. That is, a nucleus from a differentiated cell (for example, a larval intestinal cell of a frog) can be injected into a mature egg which has had its nucleus removed or destroyed, and the injected nucleus can direct normal development of a tadpole and subsequently of a fertile frog. (3) The sequences of nucleotides in the DNA of all cells in an organism appear to be the same; that is, DNA-DNA hybridization of DNA from different cell types indicates that the different cell types do not have unique DNA base sequences. Since these results indicate that all cells contain the complete genome for an organism, different cell types appear to arise as a result of the expression of unique sets of genes in each cell type. *See* CELL DIFFERENTIATION; DEVELOPMENTAL GENETICS; GENE ACTION; SOMATIC CELL GENETICS.

Determination. Initially the cells of a developing embryo are not restricted in their developmental potential or fate (that is, what type of cell they can become), but as embryogenesis proceeds, a cell's developmental potential becomes restricted or fixed. For example, after several divisions a cell from a frog embryo can develop into most frog cell types. After gastrulation, however, a surface cell in the neural plate area is restricted to develop as a neural cell type, and eventually its developmental potential will be for a specific type of neural cell. Restriction of developmental fate is called determination. Determination is an operationally defined term. In general, a cell, tissue, or organ is determined when its developmental fate is the same in its normal position in the organism, when removed from the organism and grown in isolation, and when grafted and grown at a new place on the organism. Thus, in order for determination or restriction of developmental potential to be observed, the determined state must be stable enough to withstand considerable experimental manipulation. The determined condition may however, be reversible, and the degree of stability observed varies from system to system. Determination is observed at several levels of organization (cell, tissue, organ) and occurs in both plants and animals. *See* FATE MAPS (EMBRYOLOGY).

Regulative and mosaic development. Early embryologists observed that some animal embryos divided to produce two or four cells which, when individually isolated, were capable of developing into small but normal larvae. In other species these cells were not capable of producing complete larvae but only abnormal structures which contained some of the normal cell types. The former type of embryos exhibited regulative development, while the latter exhibited mosaic development. For some time embryologists believed that these two types represented fundamentally different mechanisms of regulating development. Animals that exhibit regulative development, however, eventually exhibit mosaic development. For example, each cell of a sea urchin embryo at the four-cell stage can produce a larva, but at the eight-cell stage no single cell can undergo normal embryogenesis. Developmental biologists now believe that it is the timing of determination events which differs, and not fundamental mechanisms of development. *See* CLEAVAGE (EMBRYOLOGY); MOSAICISM.

Cytoplasmic determinants and induction. Two mechanisms have been identified that bring about determination. The first involves the presence of unique factors, called cytoplasmic determinants, which are products of the maternal genome and are located in specific areas of some animal eggs. The cells which come to contain these determinants differentiate along specific pathways. An extensively studied example is the cytoplasmic determinants in pole plasm of fruit fly (*Drosophila melanogaster*) eggs which are necessary but not sufficient for the cells to differentiate into germ cells. The second mechanism is induction, a process by which two tissues interact so that one or both differentiate along specific pathways. A classic example of induction is the action of mesoderm on the overlaying ectoderm in the frog embryo at the time of gastrulation. The mesoderm acts on the ectoderm, causing it to form the neural plate. Only ectoderm of a certain developmental age is capable of responding to the mesoderm, and this ectoderm is said to be competent. Competence or the ability to respond to an inductive signal is often a transient state. The chemical or physical base for induction and competence and the chemical identity of cytoplasmic determinants are yet to be established. *See* EMBRYONIC INDUCTION.

Molecular mechanisms. Developmental biologists have gained substantial insights into the molecular bases for determination in model organisms such as *Drosiphila*. At least three sets of cytoplasmic determinants (maternal gene products) are present in the fly egg: determinants for germ-cell formation, determinants controlling dorsal-ventral polarity, and determinants for the anterior-posterior polarity. Some of these determinants are messenger ribonucleic acids (mRNAs) coding for proteins which are transcriptional regulators (that is, proteins that regulate gene activity).

Of great interest among the genes that are activated in *Drosophila* are the homoeotic genes that specify the characteristic structures (for example, leg, wing, antenna) that are found in a particular segment of the insect. The homoeotic genes have a common, evolutionarily conserved region called the homoeobox which codes for the portion of the protein that binds to DNA and, by doing so, appears to be involved in the determination of cell fate. Homoeobox-containing genes have also been identified in plants and vertebrates. Moreover, the vertebrate *Hox* genes (vertebrate homoeobox genes) are expressed during embryogenesis appropriate for specifying cell fates. Although tremendous progress has been made in understanding cell fating in several model organisms such as *Drosophila*, there is no evidence to indicate that mammalian and plant eggs contain cytoplasmic determinants. Apparently, the initial events associated with cell fating in these groups of organisms are different from those in *Drosophila*.

Morphogenesis. Morphogenesis involves the production of form and structure by integrating the differentiation of many different cells and cell types into specific spatial patterns. This higher level of organization has been difficult to investigate in terms of establishing mechanisms. The processes of determination, competence, and induction are involved. One of the greatest challenges faced by developmental biologists is to bridge the gap between genes and patterns. It is clear that patterns are a result of gene activity, but the relationship between genes and patterns in most organisms is not well understood. *See* ANIMAL MORPHOGENESIS; PLANT MORPHOGENESIS.

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Developmental genetics

The branch of genetics which studies how genes control embryonic development. Advances in the field have emphasized the degree of conservation of the genes controlling development throughout evolution. Thus, genes which are important in such distant organisms as insects with an exoskeleton, highly segmented organization, and ventral nervous cord (*Drosophila* is the type experimental species), and vertebrates with their endoskeleton, lack of markedly segmented organization, and dorsal nervous system (the mouse is the type experimental species) share a number of highly homologous (similar) genes controlling early development. These genes are also found, and sometimes have similar roles, in descendant worms of the common ancestral annelid (although the most frequently studied descendant is a nematode, *Caenorhabditis elegans*), and some are also shared with plants.

Maternal Inheritance. Maternal inheritance is found in nematodes, mammals, and insects. For instance, many of the genes responsible for the determination of the cell's fate in the larva of *C. elegans* are expressed from messenger RNA already present in the egg; that is, they are maternal-effect genes. In this case, it is not the genotype of the zygote that influences development but that of the mother. Thus, homozygosity for a recessive mutation in the mother leads to altered development, even though the sperm is from a homozygous wild-type male and the resulting zygote was heterozygous (that is, if the organism could survive and was a female, it would lay eggs which would lead to normal development). The percentage of maternal-effect genes is also high in *Drosophila*. *See* GENE ACTION.

Imprinting. Imprinting, a developmentally important phenomenon that was first discovered in insects, is also important for mammalian development and human disease. In imprinting, genes transmitted through the testis sometimes function differently from those transmitted through the ovary. This explains why parthenogenesis does not work in mammals: the mammalian embryo requires a pronucleus derived from each sex's imprinting. Many portions of the genome have been found to be imprinted, including the reciprocal imprinting of insulinlike growth factor and its receptor. Some major human diseases occur when both a paternal and a maternal copy of a gene are not present. The Prader-Willi syndrome, a disorder of mental retardation, poor appetite regulation, and mild dysmorphic features, is an example. Advances have strongly implicated gametogenesis-specific methylation of key controlling regions in the imprinting process. Such imprints seem to be erased from the migrating germ cells en route to the developing gonad, and then are established differentially during ovigenesis and spermatogenesis, presumably by proteins uniquely expressed in the two different gonads and with specificity for the particular deoxyribonucleic acid (DNA) sequences. The actual expression of imprinting differences frequently involves competition between cis-linked genetic elements and frequently involves a nontranslated ribonucleic acid (RNA) species. *See* DEVELOPMENTAL BIOLOGY; GENETICS.

Induction. Another general phenomenon under genetic control during development is induction: the action of one cell or tissue on other cells to determine altered gene expression in them. For many years, the classic example of induction was the capacity of the dorsal blastopore lip of an amphibian embryo

to induce a second body axis after transplantation to the ventral side of another embryo; a positive-acting substance was sought. Ironically, it turns out that the signals from this organizer promote dorsal development by repressing a ventralizing signal apparently produced by a bone morphogenic protein (BMP4). Another significant inductive event, that of the notochord in patterning the neural tube, is due to the sonic hedgehog protein, one of three vertebrate hedgehog paracrine factors (that is, they affect only cells that are nearby the cell that released them).

Development control genes. Homeobox genes (*Hox* genes) are used in both insects and mammals to provide information for anterior-posterior positioning. Forkhead (forkhead box, *FOX*) genes are another highly conserved family of genes, first identified for their important developmental roles in *Drosophila*.

HOX genes. Homoeotic mutations change one paired structure to another of a more anterior or posterior compartment (for example, a leg to an antenna). The study of their structure and function has provided a paradigm for the role of genes in conveying positional information during development. In *Drosophila*, seven homoeotic genes are grouped in two complexes. Their role in establishing segmental identities is well defined, and the DNA sequence of the genes shows a highly conserved element called the homeobox. This segment codes for 60 amino acids and is highly conserved in species as distant as mammals. Importantly, this conserved sequence is also found in some pair-rule and polarity genes, and the search for genes homologous to these led to the identification of other genes that are highly conserved in animal evolution. The degree of stability of their function in evolution is reflected in the fact that the deficiencies of early development caused by mutations in a particular homeobox gene can be corrected by providing the human gene so that its product can substitute for the *Drosophila* gene.

Although *Drosophila* uses one set of homeobox genes (separated into two clusters on two different chromosomes), mammals have amplified the set of genes to a minimum of four clusters, each the size of the single cluster in *Drosophila*. These genes maintain the same patterns of expression in both mammals and *Drosophila*. They are expressed 5′ to 3′ in order of transcription, and the 5′ to 3′ order in the cluster is also reflected in the posterior-to-anterior limits of expression of the gene products.

Most mutations in *Drosophila* homeobox genes are recessive, and embryonic stem-cell knockouts have disclosed that because there is sufficient redundancy in the mammalian homeobox clusters, the homozygous absence of one homeobox gene does not always result in an apparent phenotype. In humans, dominantly inherited mutations of "gene cluster" *HOX* genes have been associated with limb and other abnormalities, whereas many dominantly inherited disorders are associated with nonclustered *HOX* genes.

Other conserved genes: FOX and Wnt. The members of the *Fox* gene family now number nearly 100. They are involved in many developmental processes. Mutations in *FOX* genes frequently cause dominantly inherited birth defects in mammals, including eye abnormalities and/or lymphedema (tissue swelling caused by dysfunction of the lymphatic vessels). A gene involved in these developmental processes, *FOXC2*, is later involved in fat cell function.

Some of these developmental control genes have a role in cancer and perform signaling functions at several developmental steps. For instance, the protein encoded by the gene *Wnt11* (a contraction of *Wingless* in *Drosophila* and *int-1*, a breast cancer gene integration site) is important in cardiac development via intracellular signaling transduction molecules (protein kinase C and *Jun* amino-terminal kinase). In contrast, another Wnt protein is involved in hair-follicle formation, in collaboration with a BMP, via an intracellular signaling cascade. The gene sonic hedgehog encodes a protein that is not only involved in neural tube induction but also in airway (lung) epithelial differentiation and in a subset of small-cell lung cancers.

Growth factors. The realization that there are many more hormonelike molecules involved in biology than was previously thought has provided for an exciting advance in organismic biology. Many of these newly discovered factors are proteins and perform roles somewhat analogous to the classic hormones of endocrinology. However, many of these molecules function in a paracrine or autocrine manner; that is, they function locally (paracrine) or on the cell (autocrine) that secretes them. The known cytokines, lymphokines, and peptide growth factors number in the many tens. Some growth-factor–like molecules remain anchored in the membrane and, therefore, must function only on neighboring cells. Other peptide growth factors adhere strongly to heparin or heparin sulfate, a common component of the extracellular matrix, and thus are limited in their diffusion and act on nearby cells. In both cases, a paracrine function can occur, whereas an autocrine function is unlikely in the former case. Nerve growth factor function is best known in the late embryo, and many of the lymphokines and cytokines are known for their roles in the adult immune and hematopoietic (blood-forming) systems. *See* CYTOKINESIS.

One stimulus to the study of the role of various growth factors in mammalian embryo development came from experiences with the culture of mammalian embryos. Fertilized embryos were early cultured to the blastocyst stage and later to the limb bud stage in the laboratory. Although the success rate was very low, this degree of success was achieved with daily changes of the media, including variations in the concentration and kind of serum supplementation, the oxygen concentration, and the degree of culture agitation. The documented need for changes in sources of serum, varying from cord blood to fetal blood, and from different species, implies that differing growth factors, which would vary between the sources of serum or plasma, might be important for embryonic development. This conclusion has been amply confirmed; and insulin, insulinlike growth factors, fibroblast and platelet-derived growth factors,

transforming growth factors, and epidermal growth factors among others, have been found to have roles in early development.

Developmental studies. Studies of embryonic development in organisms such as *C. elegans, Drosophila*, mice, and the zebrafish have shed light on the role of developmental genes and have revealed that many of the genes controlling development have been conserved throughout evolution. For instance, formation of the dorsoventral axis involves members of the transforming growth factor (TGF)-beta family of proteins [known as *decapentaplegic* in flies and BMPs in mammals], which play an important role in tissue development, cell differentiation, embryonic development, and the gene encoding the enzyme metalloproteinase (*tolloid* in *Drosophila* and BMP1 in mammals), a protein-cleaving enzyme requiring a metal ion. Other examples of evolutionary conservation include the *Pax6* gene, a member of the paired box family, which initiates the very different eye formation of *Drosophila*, of mammals, and even of cephalopods which are thought to have independently evolved a vertebrate-like eye. Although most genes involved in sex determination have not been conserved, a gene has been cloned in *C. elegans* that is highly homologous to a gene involved in the *Drosophila* sex determination cascade and to a sex determination gene in mammals (whose role has yet to be fully elucidated). These findings have resulted in the emergence of a new field of investigation known as evolutionary development, or Evo-Devo.

Caenorhabditis elegans. *Caenorhabditis elegans*, a small transparent nematode with rapid development, is highly determined, that is, a cell and its descendants are committed to particular differentiated states. In fact, each of the larva's 558 cells has an exact fate which has been carefully mapped. This fate map has been confirmed by removal of single cells, resulting in deletion of the expected descendant cells in the larva. Although the development of *C. elegans* is highly determined, there are a few examples of differentiative events that involve cell interactions, that is, where the phenotype requires cell-to-cell communication and is not a predetermined fate laid out in the egg. *See* CELL CONSTANCY; FATE MAPS (EMBRYOLOGY).

Drosophila. The complete sequence of each *Drosophila* chromosome has been obtained. Thus, almost any *Drosophila* gene that can be identified by a mutation affecting development subsequently can be cloned. In addition, the generation of new mutations by transposons (large discrete segments of DNA that are capable of moving from one chromosomal site to another in the same organism or in a different organism) generates an alternative strategy for cloning the gene. In other words, a probe for the transposon can be used to "walk" to the DNA whose expression has been altered by the transposon insertion. *See* MUTATION; TRANSPOSONS.

A powerful tool for studying *Drosophila* development involves somatic recombination techniques to create patches of homozygous mutations in an otherwise homozygous animal. These mosaics enable the study of the effects of homozygosity on a gene at a period later in development than would be possible if the mutation were homozygous lethal early in development. These studies help to readily determine if the gene effect is cell autonomous or can be corrected by diffusible factors from nearby cells; they also show that wings, legs, antennae, and eyes of the adult fly do not derive from single cells but are polyclonal in origin. In mammals, mosaicism can be studied by using embryo/embryo or embryo/embryonic stem cell chimeras. *See* CHIMERA; MOSAICISM.

Mammals. In mammals, development to the blastocyst stage occurs before implantation. The trophectoderm, primitive entoderm, and inner-cell mass have already appeared as separate lineages. In this zygote of about 100 cells, the inner-cell mass is nearly totipotent; in the mouse, as few as three inner-cell-mass cells may give rise to the adult organism. An important point about mammalian development is that it is highly regulative. In other words, the embryo has the ability to adjust its size. Bisected embryos or multiple embryos fused together lead to normal-sized fetuses. This has been confirmed for humans, where in vitro fertilization resulted in a three-eighths embryo which successfully implanted and led to a normal-sized newborn.

The long generation times of most mammals prevent them from being easily manipulated organisms for genetics. Mice, with 3 months' generation time, are the experimental organisms of choice. Humans provide much ancillary data, since physicians can screen for many natural mutations, and the large population guarantees the occurrence of many interesting natural genetic crosses. It is possible to create null mutations [changes in a gene that blocks its transcription into ribonucleic acid (RNA) and/or translation into a functional protein product] for almost any cloned gene by knocking it out (inactivating the gene) in embryonic stem cells. Alternative techniques for disrupting gene expression include the use of antisense techniques, oligonucleotides, or other stretches of nucleic acid that copy the antisense DNA strand, as well as the use of short interfering RNA (siRNA), a short sequence of RNA that finds its complementary RNA and inactivates its messenger RNA by a variety of mechanisms. *See* OLIGONUCLEOTIDE.

Zebrafish. *Danio rerio*, the zebrafish, provides a clear embryo in which the events of its early development (which mostly occur in 24 h) can be easily observed. A single female can potentially provide thousands of offspring, and homozygosity can easily be induced, which is important for studying homozygous recessive developmental genes. The generation time is not as short as that of mice, but the cost of maintaining large numbers of zebrafish is much more reasonable than that for large numbers of mice. Because of these attractive features, saturation mutagenesis for developmental pathways (that is, induction and recovery of large numbers of developmental mutations to identify all the genes affecting this function) has been undertaken in the zebrafish,

as it had previously been done in *Drosophila*. Analysis of newly found mutations has already allowed confirmation of the involvement of TGF-beta family members related to *decapentaplegic* in the dorsal ventral patterning of the zebrafish. Zebrafish have also allowed extensive genetic analysis of heart development, frequently confirming observations made in the chicken where the developing heart is more accessible. Robert P. Erickson

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Developmental psychology

The study of age-related changes in behavior from birth to death. Developmental psychologists attempt to determine the causes of such changes. Most research has concentrated on the development of children, but there is increasing interest in the elderly, and to a lesser extent in other age groups. Also, most research has focused on changes in behavior at certain ages, but there has also been work on individual differences that are constant across all ages. For example, 4-year-olds resemble each other and differ from 7-year-olds in their understanding of the world, but some individuals are more intelligent than their peers at all ages. Similarly, 1-year-olds differ from 3-year-olds in their response to strangers, but some individuals are more shy of strangers than others regardless of age. Finally, although most developmental work examines humans, there has been some work on primates and other species that would be considered unethical with human beings. Thus the sensory deprivation of kittens and the separation of monkeys from their mothers have provided information about abnormal perceptual and emotional development, respectively.

Method. Developmental psychologists who study children rely more upon careful observation in natural settings than upon laboratory experiments for several reasons. First, children react strongly to unfamiliar situations; the language capacity of young children can be judged more accurately at home with their parents than in a laboratory with strangers. Second, there are ethical reasons: it is not permissible to subject newborns to pain for research purposes, but circumcisions may be observed. Third, common sense guides the psychologists: it is not possible to require parents to train their children primarily by physical punishment or by reasoning, but families with different habits can be observed. Finally, age is the primary variable studied by developmentalists, and aging cannot be administered, but only observed.

Under these circumstances, only partial conclusions can be drawn about the causes of development. The field has been dominated by descriptive research, with increasing attempts to explain developmental phenomena by the use of animal experiments or by statistical methods.

In longitudinal research, a group of individuals is studied at regular intervals over a relatively long period of time. This contrasts with cross-sectional research, where individuals of different ages are studied at the same time. Conclusions from the two types of research may differ. For example, if intelligence is studied in different groups of people at different ages, intellectual declines may be noted as early as 20 years of age, whereas lifetime studies of a group show no decline until the age of 50 and certain functions may even improve up to that time. This is due to the fact that generations differ in average education: younger people are generally better educated than older ones. These are called cohort effects. Longitudinal research eliminates cohort effects masquerading as age-related changes. Moreover, it permits an examination of the effect of early experiences on later development in a quasicausal fashion.

Finally, case studies, that is, close and extensive observations of a few subjects, have been relied upon by important developmental theorists such as S. Freud and J. Piaget.

Stages. Most developmental psychologists believe that development occurs in stages, with dramatic differences between ages. This matter, however, is debatable. Perhaps emphasis should be placed not on stages but upon the sequences through which individuals pass at various rates. Perhaps development is so gradual that neither stages nor sequences are accurate terms; instead individuals become more and more skilled with age to a point where decline begins.

Nature–nurture issue. An explanation of developmental changes requires a judgment as to the relative importance of genetically programmed maturation and environmental influences. Although most developmentalists believe that genetic endowment and environmental experience interact to account for behavior, the degree to which either affects a particular behavior is still often debated. This issue has important implications for the success of environmental intervention in the face of genetic constraints. For example, the influence on children of parental speech versus genetic programming in language acquisition is much debated, as is the origin of gender differences in behavior. *See* BEHAVIOR GENETICS.

Theories. Developmental psychology is divided roughly between those who study personal–social (emotional) development and those who study intellectual and linguistic development, although there is a small but growing interest in the overlap between these two aspects of personality, known as social cognition. The study of personal–social development in childhood is dominated by the theory of attachment formulated by J. Bowlby and extended by M. Ainsworth. In adolescence and adulthood, E. Erikson's theory of psychosocial development is prominent. The study of intellectual development at all ages is dominated by Piaget's theory of cognitive constructivism.

Bowlby's attachment theory. Ainsworth defines attachment as "an affectional tie that one person forms to another specific person, binding them together in space, and enduring over time ... [It] is discriminating and specific." It is not present at birth, but is developed. In a word, attachment means love. Attachment behaviors such as crying, smiling, physical contact, and vocalizing are the means by which attachment is forged but are not to be equated with the more abstract, underlying construct of attachment. Attachment theory is strongly based on ethological notions. Thus, attachment is seen as serving a biological function, that is, the protection of infants by ensuring their proximity to (attached) adults. The common goal of attached individuals is proximity. Bowlby was influenced by Freud's psychoanalytic theory of development, but argues that there is a primary biological need to become attached to at least one adult, whereas Freud argued that love for a mother was secondary to her satisfaction of an infant's hunger.

Most of the work on attachment theory has explored the progressive attachment of infants to mothers, but attachment to fathers has been examined as well. Children are capable of becoming attached to other individuals, but theorists emphasize that children cared for by too many people, and no one in particular, may become attached to no one. It is also thought that the first 2 or 3 years of life are a sensitive period for the normal development of attachment. Children who do not become attached to anyone during these years are likely to experience emotional problems for the rest of their lives.

Four major phases in the development of the capacity for attachment are distinguished in early childhood (**Table 1**).

Development of specific behaviors leads to the larger changes; for example, smiling seems to be biologically programmed, since even blind infants smile. Infants smile at stimuli most likely to come from mothers, such as female voices and smiling faces. From about 3 months, infants smile more to mothers than to female strangers.

Crying is of particular interest. In the first half of the twentieth century, social learning theorists, such as J. Watson, viewed crying as a bad habit and recommended that parents ignore it. Attachment theorists, however, view crying as a signal, the infant's only means of calling the mother closer or drawing attention to a need for food. They recommend a prompt response to crying for the development of secure attachment. In one of the most important studies in the field, this dispute was resolved in favor of attachment theory: mother–infant interactions were observed over the first year of life, and it was found that the more a mother ignored her child and the longer she took to respond, the more and the longer the child cried later on.

A parent's response to a child's signals positively affects the quality of the attachment relationship between child and parent, although children become attached in some fashion even to abusive parents if there has been ample interaction. Ainsworth developed a procedure for evaluating the attachment between young children and their mothers. She observed children in an unfamiliar room responding to the presence or absence of mothers and of strangers. One group of children were clearly securely attached to their mothers, using them as a base from which to explore strange toys and people, and checking back occasionally; in their mother's absence these children protested, often trying to follow her out. Such children greeted returning mothers and sought contact with them. Other children, classified as insecurely attached, avoided contact with mothers, even turning their back upon reunion. A third group of children seemed insecure in a different way, clinging to mothers when present but refusing to be comforted upon reunion, sometimes even expressing anger. This procedure has been used in hundreds of studies, sometimes including fathers as well as mothers. Some children are securely attached to both

TABLE 1. Bowlby's theory of the development of attachment

Typical age range	Phases	Emotional characteristics, new developments
0–3 months	Undiscriminating social responsiveness	The child watches, listens to human faces; sucks and grasps to maintain contact; smiles and vocalizes more to humans than to nonhuman things; cries to attract adults and stops crying upon response.
4–6 months	Discriminating social responsiveness	The child smiles and vocalizes more to attachment figures than other humans; stops crying more readily when attended by attachment figures.
7 months–3 years	Active initiative in seeking proximity and contact	When able to crawl and walk, the child seeks out attachment figure; greets, talks to, follows, approaches, clings to attachment figure; protests figure's departure; may begin to fear strangers.
3 years on	Goal-corrected partnership	Language and cognitive development make it possible for the child to understand attachment figure's plans and to coordinate needs for proximity and contact with attachment figure's other goals.

parents, others only to one parent or to neither.

Children react strongly to prolonged separations from those to whom they have become attached. Until the age of about 3, children seem unable to understand that a parent may have good reasons for a lengthy absence. At first, they protest the parent's departure by crying, searching for them, and rejecting offers of comfort. If the separation continues, the child enters a phase of despair, crying quietly, withdrawing, and acting listless. Finally, the child recovers, seeming even cheerful, and accepting comfort from others. However, it is obvious that all is not well when the parent returns and is ignored by the child in a phase of detachment. After prolonged separations, reestablishment of trust may be slow.

Erikson's theory of psychosocial development. Like Bowlby, Erikson was a revisionist follower of Freud. Erikson deemphasized the infantile sexuality central to Freud's theory and defined psychosocial development in terms of eight crises (**Table 2**). A crisis is a conflict between two antithetical feelings. Erikson's theory agrees with Bowlby's in the sense that for him the first psychosocial crisis is the establishment of trust as against mistrust. Each age is dominated by particular psychosocial crises, and successes or failures in meeting them have lifelong consequences.

Family studies. Ever since Freud, it has been believed that early experiences, especially with the mother, are crucial in emotional development. Children have been increasingly studied in family settings, with examination of nontraditional family arrangements. Studies of the effects of maternal employment have afforded no simple conclusions, although it seems clear that day care away from the home need not cause any cognitive or emotional disturbance in children. In cases of divorce, a year of emotional crisis often follows, but subsequent disturbances are largely due to continuing parental conflict over the children.

Piaget's theory of cognitive development. Piaget was not interested in individual differences in intelligence, but rather in formulating a universal scheme of intellectual development. For Piaget, intelligence is defined as the ability to adapt to the environment, an ability that depends upon physical and psychological (cognitive) organization. Cognitive structures are termed schemata (or schemes). The adaptation process has two complementary components, assimilation and accommodation. Assimilation refers to the tendency to process new information, sometimes with distortion, in terms of existing cognitive structures. Accommodation refers to the opposite process, that is, the modification of existing cognitive structures in response to new information. In other words, whereas assimilation refers to the limits of learning, accommodation is Piaget's view of learning itself: learning occurs directly, from manipulation of materials, without external instruction or reinforcement. An infant who sucks a new rattle is assimilating it to the schema of sucking while simultaneously accommodating the sucking to the rattle: the infant has probably never sucked anything of that shape before so that the sucking must change. An individual strives for equilibrium between assimilation and accommodation, with thought being neither unrealistic (excessive assimilation) nor excessively realistic and hence disorganized (excessive accommodation). *See* COGNITION; INTELLIGENCE.

For Piaget (**Table 3**), cognition gradually becomes abstracted from perception over the course of 12 years. Cognition during at least the first 18 months is directly observable. Infants begin cognitive exploration by actively perceiving and reflexively manipulating objects, giving the name sensorimotor period to the first phase of intellectual development. Perception is a key form of early cognitive activity, especially with newborns. The newborn infant can see, hear, smell, taste, and feel much better than previously thought, though sensitivity in these areas improves throughout the first year of life.

As far as manipulation goes, Piaget termed the

TABLE 2. Erikson's theory of psychosocial development

Typical age range	Dominant crises	Outcomes of favorable ratio of positive to negative feelings
0–1 year	Basic trust vs. basic mistrust	Drive and hope; confidence that adults will meet basic needs despite temporary frustrations.
1–2 years	Autonomy vs. shame and doubt	Self-control and willpower; balanced restraint and expression despite shame and self-doubt; displays of defiance of parents.
3–5 years	Initiative vs. guilt	Direction and purpose; undertaking, planning, and executing goals despite inner guilt about some goals contemplated.
6 years–puberty	Industry vs. inferiority	Method and competence; systematic instruction produces intellectually and socially competent individual despite fears of inadequacy.
Adolescence	Identity vs. role confusion	Devotion and fidelity; a sense of self despite questions about sexual or occupational identity.
Young adulthood	Intimacy vs. isolation	Affiliation and love; commitment to partner and career despite concerns about loss of identity in this process.
Middle adulthood	Generativity vs. stagnation	Production and care; concern to establish and guide next generation at home and at work despite more self-centered concerns.
Older adulthood	Integrity vs. despair	Renunciation and wisdom; acceptance of life and death and contented reflection on a life well lived, which overcomes fear of death.

TABLE 3. Piaget's theory of cognitive development

Period and typical age range	Stages	Cognitive characteristics
Sensorimotor period (0–24 months)		
0–1 month	Stage 1: use of reflexes	Unlearned responses (reflexes) become smoother.
1–4 months	Stage 2: primary circular reactions	Repetitions of accidentally discovered sensorimotor responses that interest child; interest in sucking for sake of sucking.
4–8 months	Stage 3: secondary circular reactions	An interest in effects of actions upon objects; plays with toys.
8–12 months	Stage 4: object permanence	An ability to remember that things exist even when child does not see them; searches for things seen to disappear; appearance of genuine intentionality.
12–18 months	Stage 5: tertiary circular reactions	Experiments on objects and an interest in novelty; searches for objects seen to disappear no matter where they were found previously.
18–24 months	Stage 6: symbolic representation	Can solve problems by thinking about them rather than simply by trial and error; full object permanence in the sense that child will search for objects even when they were not seen to disappear; first abstract mental representations of objects.
Preoperational period (2–7 years)		
2–5 years	Stage 1	Deficient causal reasoning; concentration on only one dimension of reality at a time; more or less random arrangement of things in serial order; no one-to-one correspondence in (rote) counting of objects; inability to take others' perspective (egocentrism).
5–7 years	Stage 2 (transitional)	Physical causal reasoning present but primitive; vacillation about correct responses on conservation tasks; correct but laborious arrangement of things in serial order; one-to-one correspondence in counting, but length of displayed objects still misleads child; some successful role-taking but egocentric errors still occur; instruction in conservation, counting, and so on, is fruitful at this stage.
Concrete operational period (7–11 years)		Child reasons correctly; can conserve quantity, then weight, then volume, among other things because of ability to focus on several dimensions of a problem simultaneously; arranges things in serial order according to an efficient overall plan; full mastery of counting with no deception about physical distance between things counted; egocentrism has waned.
Formal operational period (11 years on)		Awareness of concrete reality as a subset of possibility; reasoning becomes more flexible and systematic; thinking is logical and in terms of abstract propositions systematically combined; mental hypotheses tested.

young child's inclination to repeat interesting activities "circular reactions." During the sensorimotor period, the child is at first interested only in its own repetitious activity (primary circular reactions), although this is an advance over simply performing reflex actions. Then the child becomes interested in actions because of the effect they have had in the past upon objects (secondary circular reactions). Finally, the child becomes interested in novel outcomes of actions on objects (tertiary circular reactions). Between the ages of 18 and 24 months, infants become capable of symbolic representation, occasionally solving problems just by thinking about them.

The major accomplishment of the sensorimotor period is object permanence, the realization that objects continue to exist even when not observable. At 8 months, the infant first demonstrates some such capacity, but it is not until the end of this period that an infant can find an object that was not observed to disappear.

During the next 5 years, sometimes termed the preoperational period, children work on concrete operations such as classifying objects into categories (animals, furniture), arranging things in serial order, figuring out causes and effects, or understanding a one-to-one correspondence of numbers to objects counted. They also eventually manipulate reality enough to overcome perceptual illusions such as that

an amount of water changes when it is poured from a short wide glass into a tall narrow glass. In other words, they grasp the notion of conservation of quantity. From 7 to 11 years, children further consolidate their concrete mental operations, grasping conservation of volume, for example, only at the age of 11. The phenomenon that conservation of different aspects of reality is understood at different ages is called horizontal decalage.

At about 12 years, many adolescents enter the final stage of intellectual development: formal operations. They become capable of abstract, logical thought. They understand reality as a subset of possible worlds, and are able to form multiple, systematic hypotheses, involving all possible combinations of relevant variables, in order to explain things.

An important point for Piaget is that children actively construct reality for themselves; they need not, indeed cannot, be instructed in many basic developments. Another of his concerns was the formulation of a universal scheme of intellectual development. Many quarrel with his age assessments of children. He underestimated the capabilities of young children because he used tasks and materials unfamiliar at the age tested. Others argue that his stages are not as coherent in real children as his theory suggests. But most people accept his sequence of stages as useful for classifying children.

The most important criticism of Piaget is directed at his concept of egocentrism. He argued that infants are born incapable of distinguishing themselves from reality. Even preoperational children (2 to 7 years) were said to be so egocentric as to be incapable of realizing that others view the world differently from themselves. Egocentrism was a cornerstone of Piaget's theory, the cause of all other deficiencies of thought in the young child. As evidence of egocentrism, Piaget pointed to the following: children often do not specify pronouns properly, misuse causal language, and talk in front of others without successful communication. However, many investigators have found that children as young as 3 can grasp the perspective of others. Children may appear egocentric, but this is a result, rather than a cause, of a lack of language skill, for example, they use a pronoun when a noun is unknown. Furthermore, it has been argued that speaking out loud without regard for communicating with anyone often serves the function of self-regulation ("I want to draw something. First I need a crayon and a piece of paper."). Children plan their actions by thinking out loud, a finding that has been supported by much subsequent research. In short, language is instrumental in developing thought. Vygotsky's realization that intrapersonal processes (such as thought) are transformed interpersonal relations (such as conversation) has reestablished a meaningful role for the instruction of children, which was overemphasized by behaviorists and subsequently underemphasized by Piaget.

Language development. There is less common ground than there perhaps should be between students of developing thought and those of developing language. The study of developing language began in a serious way when N. Chomsky rejected Skinner's behavioristic notions that language was learned in the same way that rats learn to press a bar. Chomsky argued that children are born with an innate ability to acquire language. His theory of grammar paved the way for descriptions of how language structures develop. Vocalizations at birth are limited to crying, but are followed by cooing within a few months. At 6 months, most babies begin babbling, uttering their first word at about 1 year. They speak one word at a time until the latter half of the second year, when vocabulary explodes and words are combined two at a time. These early combinations are components of full sentences. Over the next few years, children increase the length, coherence, and complexity of their utterances, telling a good personal narrative by the age of 6. In the early school years, children learn to read, decoding words and simple stories and becoming more expert. Finally, from later elementary school on, children develop writing ability. Although these are the central language accomplishments at various ages, the development of each component starts earlier and ends later than the age range specified. Thus, preschool children engage in pretend reading and invent spellings for words, and elementary school children continue to correct their pronunciations.

Children acquire language in the context of language games played with parents, other adults, and children. Adults are more repetitious and use simpler sentences when talking with children than with adults, so that children pay more attention. *See* PSYCHOLINGUISTICS.

Memory development. Language and thought are fairly well developed by age 12. Once people have formal operations and are capable of writing coherently, they continue to operate at this level throughout life. But the issue of memory deterioration with age has considerable importance. There are many types of memory, but the most important developmental change in old age recalls the most important developmental change in childhood, that is, the use of memory strategies. As children progress through elementary school, they repeat to themselves things they wish to remember and organize information they wish to learn, perhaps forming an image to link such items as a state and its capital; such strategies improve memory. In old age, the reverse seems to happen: memory strategies are used less and memory suffers. Instruction in memory strategies helps children and the elderly. *See* MEMORY.

Kohlberg's theory of moral development. L. Kohlberg's work on moral development spans the chasm between intellectual and emotional development. He studied reasoning about hypothetical moral dilemmas, such as whether a person should steal an unaffordable drug in order to save someone's life. He classified such reasoning in six stages, two at each of three levels of moral development. At birth children are considered to be premoral. By the age of 7, most children are in stage 1, chiefly characterized by the belief that people should act in certain ways in order to avoid physical or other punishment. In 2 or 3 years, children reason primarily in terms of doing things for rewards; this is stage 2. These first stages

are termed preconventional morality; rules are not yet internalized.

At the conventional stage of moral reasoning, rules are internalized. Stage 3 of this period is often termed good-boy morality, and involves reasoning focused less on rewards than on maintaining the approval of others. Stage 4 involves reasoning that unquestioningly accepts conventional rules. Actions are judged by a rigid set of regulations, religious, legal, or both. Most individuals do not develop past this point.

A few, however, do reach postconventional moral reasoning, stage 5. These individuals think in terms of moral principles. Law is viewed as a social contract, and if it violates the rights of even a few, it is considered unfair and in need of change. Rarely, a step higher to stage 6 is reached, governed by original abstract moral principles such as articulation of the golden rule.

Moral development occurs through exposure to advanced thinking. Kohlberg argued that moral development is progressive, without regression to earlier stages. Although he interviewed only boys, he considered his scheme as universal and received some support. It has, however, been argued that women are too often assigned to stage 3 for life because of an overriding concern with seeking approval and not offending others. According to this analysis, female moral development is oriented toward an equally compelling ethic of care that is misunderstood and underestimated in male standards of justice. However, in most studies using Kohlberg's system, there are no significant differences between large numbers of women and men.

Gender differences and sex roles. Only a few differences in behavior have consistently been found between boys and girls. These are small and apply only to the "average" boy and girl. Individual boys differ from each other much more than from girls. On the average, boys are more aggressive than girls and slightly better at mathematics and spatial relationships. Girls, however, are more verbal than boys.

Most of the differences between boys and girls lie in the sex roles they construct for themselves. Before they are 3, children know their own sex, but it takes another year or two to realize that sex is fixed for life. Before the age of 5, boys may not realize that their sex does not change even if they played with dolls. Preschool children prefer toys traditional for their sex. The causes of the observed differences have been strongly debated.

Other differences. Some cultures apparently express formal operations in a way very different from that proposed by Piaget, or not at all. Generational differences extend to education. Individual differences, as in temperament, are relatively stable over a lifetime. It is thus debatable whether any universal scheme of development is possible.

Developmental psychopathology. Some individual differences are acceptable (for example, high versus low activity), but others are not. Traditionally, child clinical psychology (abnormal development) and the study of normal development were separate. However, effort is being made to integrate them. Normal developmental milestones define abnormal development; abnormal development is informative about normal processes. The serious disorders of childhood include autism, attention-deficit disorder with hyperactivity, and depression. Viewed another way, abnormal children are either overcontrolled (obsessive-compulsive) or undercontrolled (impulsive, aggressive). Developmental psychopathologists, however, are interested not just in disordered development in childhood, but in abnormal individuals over their lifetime. Such studies can shed light on the effectiveness of treatments and on the way in which disorders such as hyperactivity may be displayed differently at different ages.

Applied developmental psychology. The best work in developmental psychology has always had practical implications. Nevertheless, there is increasing emphasis on making research findings available to the public and on taking an active role in developing policy. Further applied research can benefit many groups.

Allyssa McCabe

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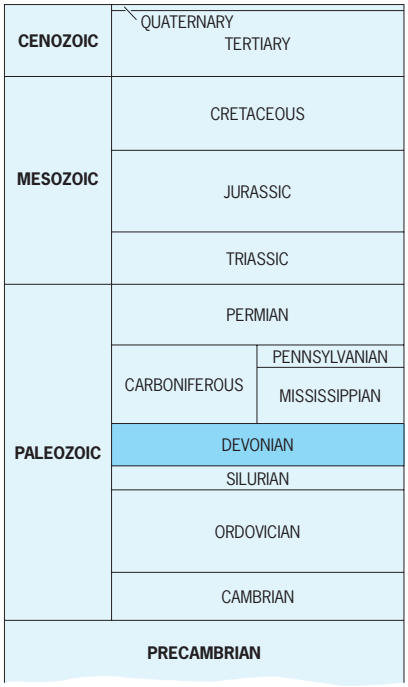
Devonian

The fourth period of the Paleozoic Era, encompassing an interval of geologic time between 418 and 362 million years before present based on radiometric data. The Devonian System encompasses all sedimentary rocks deposited, and all igneous and metamorphic rocks formed, during the Devonian Period. It is conventional that recognition of Devonian time



is determined by the definition of Devonian rocks. *See* PALEOZOIC.

The base of the Devonian System has been fixed, by international agreement, at an actual outcrop of sedimentary rocks at Klonk in the Czech Republic, where it corresponds to the base of the *Monograptus uniformis* graptolite zone. The top of the Devonian System, corresponding to the base of the Carboniferous System, was similarly fixed at LaSerre in southern France, recognized by the base of the *Siphonodella sulcata* conodont zone.

In 1839 Adam Sedgwick and R. I. Murchison proposed the Devonian System to encompass the marine sedimentary sequence between the Silurian and Carboniferous in the counties of Cornwall and Devon in southwestern England. In early-studied locales such as Wales and Scotland, only terrestrial Old Red Sandstone lies above the Silurian. When fossil corals found in marine rocks in Devonshire were considered by William Lonsdale to be intermediate in character between those of the better-known Silurian and Carboniferous marine deposits, these Devonshire rocks and the marine fossils they included were determined to be in the stratigraphic position of the Old Red Sandstone. This stage-of-evolution judgment was made in 1837, before Darwin's theory of natural selection had appeared. Because the rocks and faunas of the structurally complex type region were poorly known, Murchison and Sedgwick traveled to Germany in 1839 to see if their new system could be identified in the richly fossiliferous sequences below the Carboniferous in the Rhine Valley. Subsequently the German and Belgian Devonian sequences have served as worldwide standards of comparison.

About the same time, work in America by the paleontologists James Hall, John Clarke, and others was making known the superb physical and faunal development of Devonian rocks in New York, where there are few structural complications. If the Devonian had not been defined until 1842, it might now be called "Erian" after the New York deposits, which were described in much detail in the early reports of the New York State Museum of Natural History.

In modern times, it has become evident that even the Rhenish and New York sequences represent the Devonian System less than adequately. These sequences have been supplemented by the complete Lower Devonian sequence in the Barrandian region of the Czech Republic, and by other Devonian sequences in Iowa, Nevada, Australia, Canada, China, Morocco, and Siberia.

Subdivisions. The Devonian is customarily divided into Lower, Middle, and Upper series and their corresponding epochs. These, in turn, have been divided into stages and their corresponding ages (**Fig. 1**). The base of the Middle Devonian is at the base of the German Eifelian, or base of the *partitus* conodont zone. The base of the Upper Devonian is near the base of the Belgian Frasnian, or base of the former lower *asymmetricus* conodont zone (MN Zone 1), defined by the first occurrence of *Ancyrodella rotundiloba*. *See* CONODONT.

Determination of time equivalency (correlation) and age dating in the Devonian are usually accomplished biostratigraphically, by utilizing zones based on the evolutionary successions in individual fossil groups, or on a composite zonal framework based on several fossil groups.

A standard biochronology for the Devonian was first developed by utilizing ammonoid cephalopods, and this was most successful in the Upper Devonian. During the 1960s and 1970s, a standard biochronology was developed for most of the Lower Devonian, based on graptolites. Many local correlations within the Devonian have depended upon the faunal succession of other animal groups (for example, brachiopods, corals, ostracods, trilobites, dacryoconarid tentaculites, fishes) or on plant fossils. A microfossil zonal biochronology has been developed for conodonts to a point where it is superior to any other for the Devonian, in terms of both precision and wide applicability (Fig. 1).

Lithofacies and paleogeography. Because certain sedimentary rock types form under limited ranges of climatic conditions, mapping the nature and distribution of these sedimentary rock types provides a view of ancient climatic distribution. Climate is controlled by intensity of solar insolation and its effect on atmospheric circulation, which is related directly to latitude and direction of Earth's rotation as modified by position of landmasses. The abundance of tropical-climate rocks in present northern temperate continents, combined with lack of these tropical-climate rocks in large portions of southern tropical continents during much of the Paleozoic, has constituted one of the lines of evidence supporting the concept of continental drift. This is particularly obvious in the Devonian, and a recent reconstruction of continental positions has accommodated these climatic inconsistencies in an actualistic view of the Devonian world (**Fig. 2**), which also explains the heretofore enigmatic faunal distribution patterns. *See* PALEOGEOGRAPHY.

North America. Devonian rocks across most of the interior of North America, from Hudson Bay to the Ohio Valley across to Nevada and north to the Mackenzie Basin, consist of a predominantly marine carbonate suite, including both limestone and dolomite, commonly with abundant fossils, but containing relatively minor amounts of sandstone and shale. Interbedded with the carbonates in several areas are evaporites ranging from least soluble sulfates, the most common and widespread, through halite, common and widespread in western Canada and present in the Hudson Bay area and Michigan, to the most soluble potash salts, which are rare among evaporites of all ages, but which form a thick local sequence that is mined in the Canadian province of Saskatchewan. This carbonate-evaporite suite indicates a dry tropical climate for central North America in the Devonian. Shale- and sandstone-rich sequences characterize two belts in the North American Devonian. One extends along the Pacific coast from California through the Yukon to the Canadian Arctic islands, and probably represents the equatorial rainfall belt. The other extends along the Appalachians, and probably represents the southern warm temperate rainfall belt. This belt in

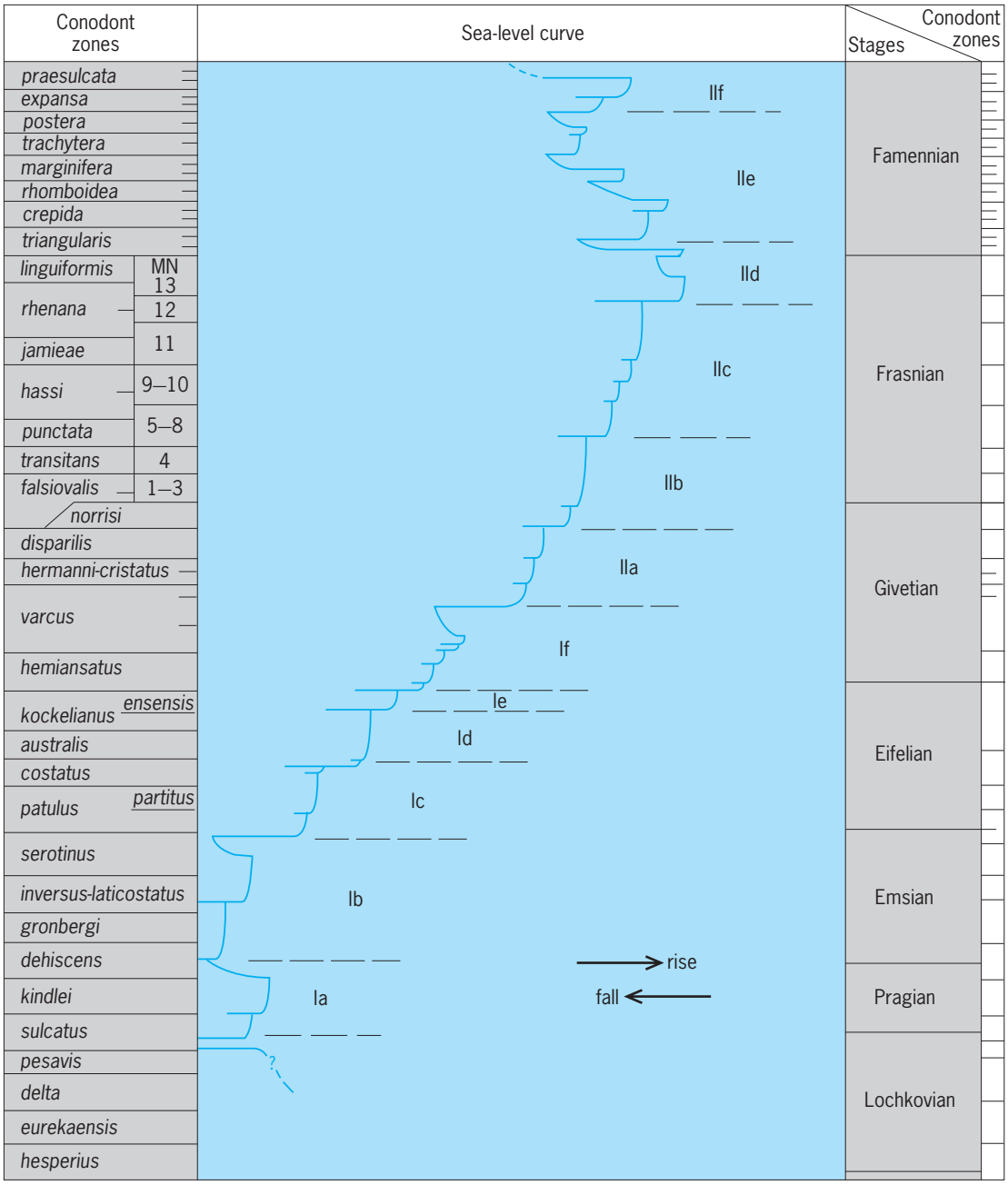


Fig. 1. Biostratigraphic zones based on conodonts (left) plotted to show approximate correspondence to the chronostratigraphic stages and series which compose the Devonian System (right) and relative sea-level changes (center) showing major fluctuations during two depositional phases (roman numerals).

New York and Pennsylvania includes widespread black shales and the famous Catskill deltaic complex (clastic wedge) that figured strongly in the early history of Devonian investigation in North America.

Europe. During the Devonian, Europe was apparently joined to North America as a result of the Silurian Caledonian orogeny to form a larger continent known variously as Euramerica, Laurussia, or the Old Red Continent. The last name derives from the Old Red Sandstone, among the first Devonian formations described, which is the coarse nonmarine detrital deposit occurring throughout most of Britain, Scandinavia, Spitsbergen, and eastern Greenland and giving evidence of the Caledonian mountains along the west. A dry tropical carbonate-evaporite sequence covers much of Russia, with equatorial bauxite-bearing thick carbonate and detrital sequences along the Ural mountain chain, the site of a seaway during the Devonian. Sequences of detrital rocks and tropical reefy carbonates characterize central Europe, particularly Germany and Belgium, with possibly deeper-water deposits in parts of southern Europe.

Asia. The largest modern continent apparently consisted of as many as 11 microcontinents during the Devonian. The largest of these was Siberia, which has dry tropical carbonate-evaporite sequences and redbeds across much of its area and more humid detrital and reefy carbonates along certain edges. The other Asian fragments also have warm-climate carbonate and detrital sequences, with equatorial

bauxites in southern China and Kazakhstan, and evaporites in Iran and other parts of Kazakhstan.

Gondwana. During much of the Paleozoic this giant continent consisted of Africa, South America, Antarctica, Australia, India, and Arabia fitted together, with smaller fragments such as Madagascar, like pieces of a jigsaw puzzle. Over most of Gondwana, including all of South America, sub-Saharan Africa, and Antarctica, Devonian rocks consist entirely of nonred sandstone and shale sequences. These indicate a humid climate, and the absence of any warm-climate rocks such as carbonates, evaporites, or bauxites strongly suggests a cool climate. Thus Gondwana was apparently centered near the South Pole in the Devonian, with most of it in the cold temperate humid belt. Only around the outer fringes in the warm temperate belt do carbonate rocks appear, in northwestern Africa, the northern Indian subcontinent, and most of Australia. Evaporites in parts of Australia indicate that this part of Gondwana extended farthest from the pole, into the dry tropical zone.

Biogeography and ocean currents. Devonian fossils are found to be distributed in three realms. Within each realm there is taxonomic similarity, which indicates that there was reproductive interchange among members of the same phyletic groups, but between each two realms there are various degrees of taxonomic dissimilarity, which indicates that there were various degrees of reproductive isolation among members of the same phylogenetic groups.

The largest realm covered Australia, Asia, Europe, western North America, and the Morocco-India fringe of Gondwana, and is termed the Old World Realm. It was unified by relatively free flowage of the warm equatorial currents and their immediate branches among the continental masses throughout this tropical to subtropical region (Fig. 2).

The Appalachian Realm covered most of eastern North America and the Colombia-Venezuela-Amazon part of northern South American Gondwana, which was adjacent to Appalachian North America during the Devonian. This region was bathed by the temperate southern west-wind current, which crossed a sufficiently broad stretch of ocean so that many Old World larvae could not make the journey, allowing endemic Appalachian forms to develop locally.

The Malvinokaffric Realm covered central Gondwana, including southern South America, southern Africa, and Antarctica. Absence of certain carbonate-secreting groups such as stromatoporoids and green algae, and great reduction of others such as corals suggest that this was a cold-water fauna, which accords with the position of this part of Gondwana suggested by the exclusively detrital rock suite. This region was bathed by a small subpolar ocean current, which was derived from the temperate west-wind

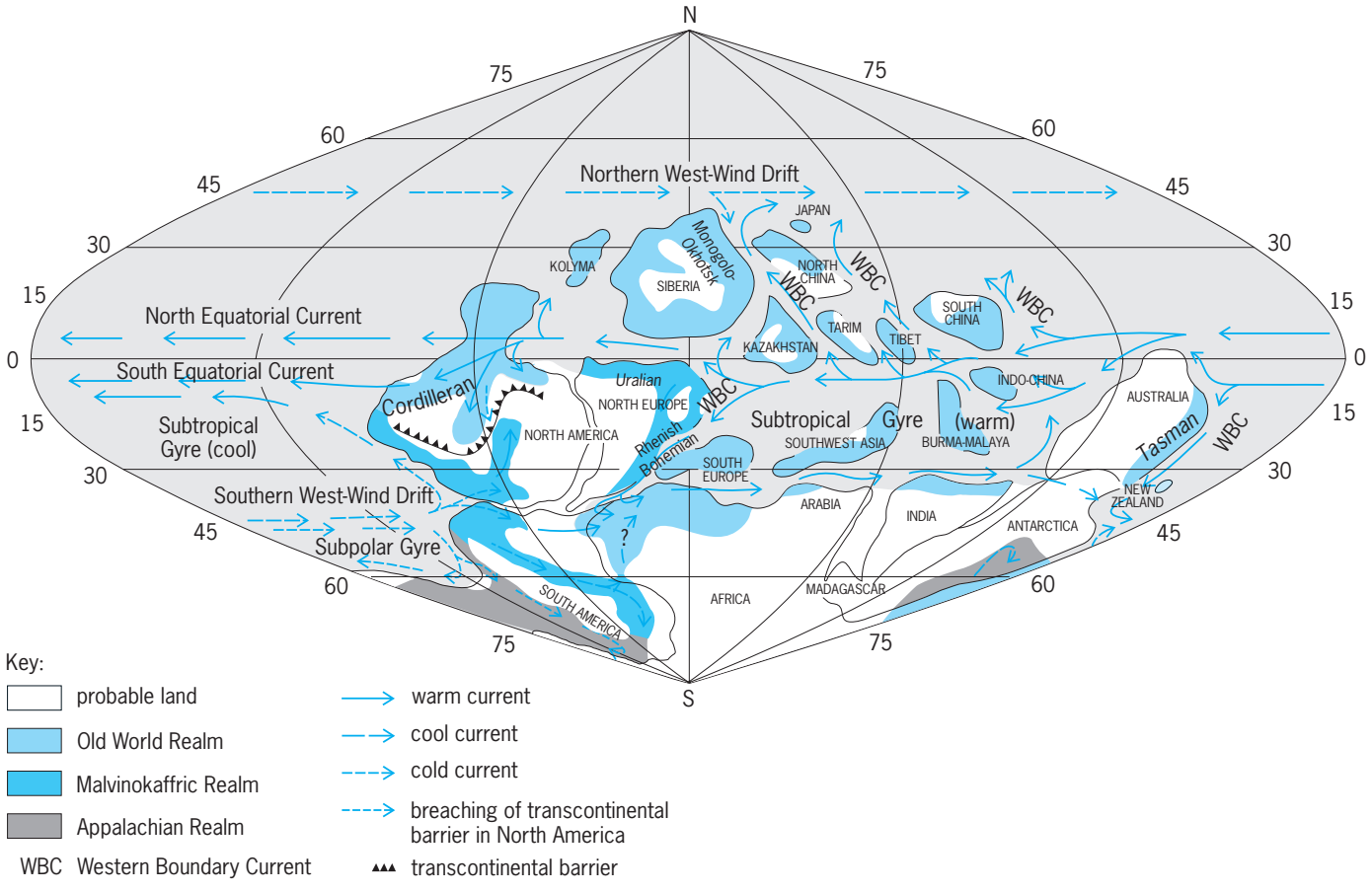


Fig. 2. Reconstruction of Middle Devonian continental positions based on distribution of lithologic climatic indicators. Probable ocean currents derived from this paleogeography explain the biogeographic realms. Kolyma and Tarim are microcontinents of the devonian. (*After M. R. Hourse, C. T. Scrutton, and M. G. Bassett, eds., The Devonian System, Palaeontol. Ass. Int. Symp. Spec. Pap. 23, pp. 99–123, 1979*)

current, but in which colder temperatures provided a barrier to Appalachian organisms.

Tectonics. Devonian mountain building was particularly noticeable along the margins of Euramerica. The Acadian orogeny formed mountainous highlands accompanied by a chain of granitic intrusions from Nova Scotia to Pennsylvania during much of Devonian time. These mountains formed a barrier that prevented mixing between organisms of the Appalachian Realm and those of the Old World Realm at the same latitude in central Europe. Erosion from the Acadian mountains produced the thick Catskill deltaic complex of New York and Pennsylvania, which spread its fine-grained sediments far into the interior of eastern North America during later Devonian time. Simultaneously, along the Arctic margin of Canada, the Ellesmerian orogeny was producing folded mountains whose erosional products formed a clastic wedge that was a mirror image of the Catskill deposits.

During latest Devonian time, the Roberts Mountains thrust, of the Antler orogeny in Nevada and Idaho, formed at the top of a subduction zone along which continental crust and overlying sediments were descending beneath oceanic sediments. A Late Devonian orogeny also affected eastern Australia. *See* OROGENY; PLATE TECTONICS.

Sea-level changes. Times of active plate movement, such as the Middle and Late Devonian, were times during which the oceanic rise systems formed very large submarine mountain chains, like the Mid-Atlantic Ridge of today. Expansion of these rise systems reduced the volumetric capacity of the ocean basins, so that marine waters rose eustatically and spread as transgressions across low-lying continental platforms, forming broad epeiric seas. This not only affected greatly the physical surface of the Devonian world, but was a major environmental factor in the evolution of plants and animals. During the earliest Devonian, the world's oceans were eustatically low, so that epeiric seas were absent from continental interiors, a continuation from the regression during the Late Silurian. Offshore marine organisms of the earliest Devonian had much in common with their predecessors.

Epeiric seas then expanded sporadically during the remainder of the Early Devonian. This transgressive episode affected Gondwana, in addition to the Euramerican platforms. The general transgressive trend greatly increased near the beginning of the Middle Devonian (transgressive part of transgressive-regressive cycle Ic, Fig. 1), and culminated, after a series of transgressive-regressive cycles, at the end of the Frasnian Age of the Late Devonian. During latest Devonian time, epeiric seas were extensive, but less evenly continuous than before orogenic movements had begun to modify the crustal surface. The interior of Gondwana did not undergo transgression after the early Middle Devonian.

As the Devonian transgressions progressed, many offshore marine animals adapted themselves to the expanding habitats of epeiric seas and then migrated widely as barriers were inundated. This led to loss of isolation, rise of competition, lowering of overall diversity, and loss of the separate realms, with replacement by a cosmopolitan Frasnian marine fauna derived from the Old World Realm. At the end of the Frasnian, still unexplained extinctions of many marine groups further reduced the organic diversity of the Devonian world.

Life. Among the marine invertebrates, trilobites (Arthropoda) were much less abundant than during the Cambrian. The planktic members of the extinct graptolites died out during the Early Devonian, at about the same time as the pelagic ammonoid cephalopods first evolved. The externally two-shelled brachiopods were at their greatest diversity, being represented by more than 900 genera. Lime-secreting corals and stromatoporoids were important and widespread in warm-water environments, and formed reefs during the Middle and Late Devonian. The extinct microfossil group known as conodonts was abundant, widespread, and rapidly evolving during the Devonian, so that conodont fossils are now regarded as the principal tools to be used for international correlation and relative age determination (Fig. 1). *See* BRACHIOPODA; CONODONT; GRAPTOLITHINA; MICROPALEONTOLOGY; STROMATOPOROIDEA; TRILOBITA.

The great diversification and radiation of fish in the Devonian has led to the term "Age of Fishes" for the period. Placoderm fish, among the most primitive of the jawed vertebrates, were successful predators in Devonian waters, and some grew to lengths up to 8 m (25 ft) just before their extinction at the end of the Devonian. The sharks, with a cartilaginous skeleton but lacking a swimbladder, may have evolved from an early placoderm. *See* PLACODERMI.

Bony fishes or Osteichthyes, a class that includes all modern fish other than sharks and agnathans, were represented in the Devonian by the primitive acanthodians, but more modern groups of bony fishes appeared in the Early Devonian. Lobe-finned bony fishes, or sarcopterygians, include both the lungfish (Dipnoi) and crossopterygians in the Devonian. The oldest known amphibians, including *Acanthostega* and *Ichthyostega*, which evolved from the rhipidistian crossopterygians, occur in strata thought to be high Upper Devonian. *See* DIPNOI; OSTEICHTHYES; SARCOPTERYGII.

Many Lower and Middle Devonian fish fossils are now known from rocks deposited in open marine environments, indicating that their habitat was marine as well as estuarine or fresh water. The vertebrates appear to have made the complete transition from ocean to dry land within the Devonian, perhaps as a result of the evolution of land plants during that period.

Land plants began to flourish near the beginning of Devonian time, and were exemplified by the vascular genus *Psilophyton* of the phylum Psilopsida. The latter gave rise in the Devonian to the Lycopsida (scale trees) and Pteropsida (true ferns). The pteropsids remain important in the world flora of today. *See* LYCOPHYTA; PALEOBOTANY; PSILOTOPHYTA; PTEROPSIDA.

J. G. Johnson; P. H. Heckel; D. J. Over

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Dew

The deposit of liquid water resulting from condensation of atmospheric water vapor to exposed surfaces that cool during the night. Dewfall is noticeable in the early morning after a calm, cool, clear night, usually as beads of liquid water on the outside and upward-facing surfaces of trees, buildings, and so forth. If the ground is moist, some of the condensed water can be the result of distillation of surface moisture. Dew forms when the surface temperature drops sufficiently to saturate air in contact with the surface (that is, when the surface drops to below atmospheric dew-point temperature); when the surface cools to below freezing temperature, frost occurs. *See* DEW POINT; FROST.

When water is plentiful, vegetation sometimes exudes liquid water through openings in the skin of leaves. This phenomenon, known as guttation, also contributes to the accumulation of liquid water on foliage, but is unrelated to the condensation mechanisms that cause dewfall. Guttation is often recognizable as droplets of water at the tips of leaves.

Dewfall is difficult to measure because of the small amounts of water involved, but absorbent paper is sometimes used to soak up and weigh the condensation on sample surfaces. In some studies, sensors have been attached to foliage to detect the formation of liquid water. Dewfall rates have been computed by using the surface heat energy budget: the latent heat associated with the condensation is measured, and the average rate of liquid water deposition is then calculated. The rate at which dew is deposited is limited by the surface energy budget. Nighttime cooling rates are greatest when the atmosphere is driest because radiative cooling is more rapid. Also, since clouds can limit infrared radiation loss from the surface, dewfall tends to be a clear-sky phenomenon.

Hygroscopic particles on surfaces can act as sites for condensation at temperatures higher than the atmospheric dew-point temperature. Thus, if a surface is not clean, the first deposit of moisture from the air can occur well before the surface cools to dew-point temperature. For some chemicals, such as common salt, condensation can start to occur when the local relative humidity reaches 80%; the humidity must be 100% in the case of a clean surface. Some desert plants exude hygroscopic salts from the interior of leaves which provide preferred sites for condensation and thereby create a supply of water for the plant. *See* HUMIDITY; HYDROLOGY; PRECIPITATION (METEOROLOGY); VAPOR PRESSURE. Bruce Hicks

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Dew point

The temperature at which air becomes saturated when cooled without addition of moisture or change of pressure. Any further cooling causes condensation; fog and dew are formed in this way.

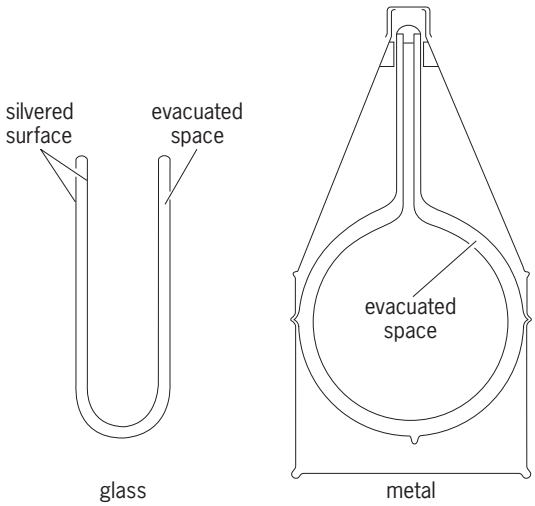
Frost point is the corresponding temperature of saturation with respect to ice. At temperatures below freezing, both frost point and dew point may be defined because water is often liquid (especially in clouds) at temperatures well below freezing; at freezing (more exactly, at the triple point, +.01°C or 32.02°F) they are the same, but below freezing the frost point is higher. For example, if the dew point is −9°C (16°F), the frost point is −8°C (17.5°F). Both dew point and frost point are single-valued functions of vapor pressure.

Determination of dew point (or frost point) can be made directly by cooling a flat polished metal surface until it becomes clouded with a film of water or ice; the dew point is the temperature at which the film appears. In practice, the dew point is usually computed from simultaneous readings of wet- and dry-bulb thermometers. *See* DEW; EVAPORATION; FOG; HUMIDITY; HYGROMETER; PSYCHROMETER; VAPOR PRESSURE. J. R. Fulks

Dewar flask

A vessel having double walls, the space between being evacuated and the surfaces facing the vacuum being heat reflective. It was invented in 1892 by James Dewar as a container for liquid oxygen.

Dewar's original flasks were made of glass with a coating of mirror silver; this type is still used in the laboratory. But for shipment and storage of liquid gases, metal vacuum vessels are used (see **illus.**). Metal vessels with a capacity of 13 gals (50 liters) can preserve liquid oxygen with an evaporation loss of only 4% per day. Evaporation rates in 29,000-gal (110,000-liter) vessels designed for transport by railway are approximately 0.1% per day for oxygen and 0.8% per day for hydrogen.



Two typical Dewar containers.

Thermos Bottle is a trademark for a Dewar vessel for hot and cold foods. H. W. Russell; G. R. Harrison

Dewaxing of petroleum

The process of separating hydrocarbons which solidify readily (waxes) from petroleum fractions. Removal of wax is usually necessary to produce lubricating oil which will remain fluid down to the lowest temperature of use. It is therefore an important step in the manufacture of lubricating oils. The wax removed may be purified further to produce commercial paraffin or microcrystalline waxes.

Most commercial dewaxing processes utilize solvent dilution, chilling to crystallize the wax, and filtration. The MEK process (methyl ethyl ketone–toluene solvent) is widely used. Wax crystals are formed by chilling through the walls of scraped surface chillers, and wax is separated from the resultant wax-oil-solvent slurry by using fully enclosed rotary vacuum filters. In one process modification, most of the chilling is accomplished by multistage injection of very cold solvent into the waxy oil with vigorous agitation, resulting in more uniform and compact wax crystals which filter faster.

In the propane process, part of the propane diluent is allowed to evaporate by reducing pressure, so as to chill the slurry to the desired filtration temperature, and rotary pressure filters are employed.

Other solvents in commercial use for dewaxing include MEK-MIBK (methyl isobutyl ketone), acetone-benzene, dichloroethane-methylene dichloride, and propylene-acetone.

Older dewaxing processes are centrifugal dewaxing, applicable only to heavy residual stocks, utilizing naphtha dilution, indirect chilling, and centrifugal separation; and cold pressing, applicable only to low-viscosity light lube fractions, in which the crystallized wax is separated from the chilled, undiluted oil in plate-and-frame-type pressure filters.

Complex dewaxing requires no refrigeration, but depends upon the formation of a solid urea-*n*-paraffin complex which is separated by filtration and then decomposed. This process is used, to a limited extent, to make low-viscosity lubricants which must remain fluid at very low temperatures (refrigeration, transformer, and hydraulic oils). Similar use is anticipated for the catalytic dewaxing process, which is based on selective hydrocracking of the normal paraffins; it uses a molecular sieve-based catalyst in which the active hydrocracking sites are accessible only to the paraffin molecules. *See* PETROLEUM; PETROLEUM PROCESSING AND REFINING; WAX, PETROLEUM.

Stephen F. Perry

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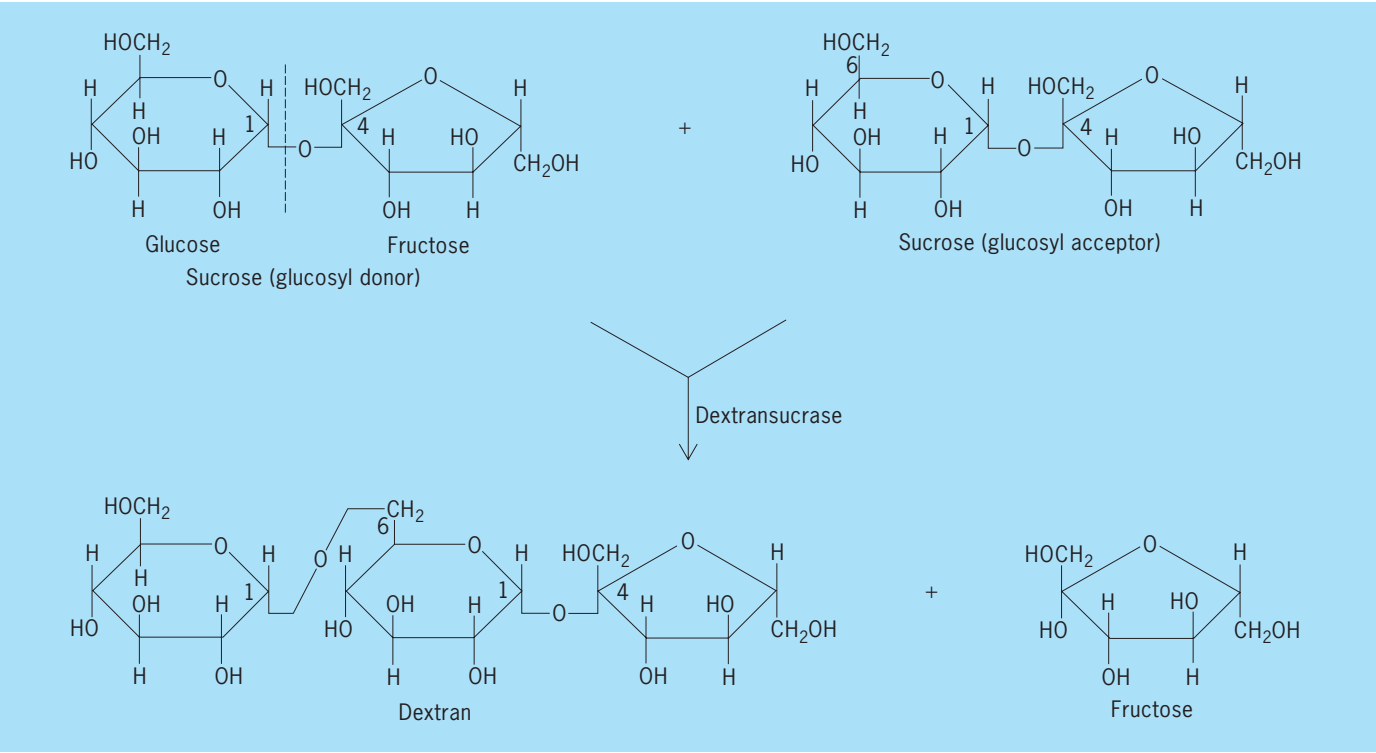
Dextran

A polyglucose biopolymer characterized by preponderance of α-1,6 linkage, and generally produced by enzymes from certain strains of *Leuconostoc* or *Streptococcus*. The chemical properties may be modified chemically for specific usages. While formerly its principal utility was as blood plasma substitute, dextran is also employed as packing material in column chromatography and as pharmaceutical agents; its average molecular weight determines usage to a great extent. Dextran's chemical and physical properties depend upon the strain of microorganism employed and the environmental conditions imposed upon the bacterium during growth, or the reaction conditions where an enzymatic method of dextran production is employed. *Leuconostoc* and *Streptococcus* species primarily convert sucrose to dextran and fructose. *Acetobacter* species convert dextrin to dextran; the α-1,4 linkage is converted to an α-1,6 linkage. *See* CHROMATOGRAPHY; POLYSACCHARIDE.

The conversion of sucrose to dextran and its by-product fructose is a transglucosylation reaction. This transformation is mediated by the enzyme dextransucrase. It is readily obtained extracellularly from suitable strains of *L. mesenteroides* which are propagated under appropriate conditions. *Leuconostoc mesenteroides* NRRL B-512F has been adopted by practically all of the Western nations for the production of dextran.

Dextran, like glycogen and amylopectin, is a branched glucose polymer but differs from amylopectin in that the principal linkage between the anhydroglucose units is of the α-1,6 type; the others are of the 1,4 or 1,3 type. The enzyme system from *L. mesenteroides* NRRL B-512F forms a polymer with about 95% α-1,6 type of linkage. The ratio between the 1,6 and the non-1,6 linkages may vary from 20:1 to almost 1:1, depending on the bacterial strain, and is characteristic of the enzyme obtained from it.

Mechanism. Formation of dextran from sucrose involves the reaction shown in the **illustration**.



Reaction for formation of dextran from sucrose.

The reaction requires two substrates. Dextransucrase is more specific in its requirement for the glucosyl donor substrate (sucrose) than for its acceptor substrates. The acceptor substrate may be sucrose, maltose, isomaltose, α-methylglucoside, and low-molecular-weight dextran. The average molecular weight of the dextran polymerized is controlled by both the type and quantity of the acceptor substrate incorporated either in the fermentation medium or in the enzymatic reaction.

The kinetic constants can be readily varied since dextran is formed from a system where double substrate kinetics apply. Neither the V nor the Michaelis constant for the polymerization reaction is fixed; they are varied by alteration in the concentration of either cosubstrate (glucosyl donor and glucosyl acceptor). In some respects, the mechanism of the polymerization is similar to that of the condensation type in that water is eliminated. It also displays, however, certain characteristics of the chain-reaction type in that the molecular weight of the dextran produced is essentially that which is desired and requires no further modification. *See* ENZYME.

Production processes. The average molecular weight of dextran produced by most of these organisms is generally on the order of several to hundreds of millions. One strain of *Streptococcus* produces, however, a dextran with a molecular weight of about 100,000 or slightly lower. The high-molecular-weight dextran, as ordinarily produced by *L. mesenteroides,* is hydrolyzed to a product with an average molecular weight of about 75,000 for use as a blood plasma volume expander. More information is available on the utility of dextran from *L. mesenteroides* NRRL B-512F for use as blood plasma substitute than on the dextran from the *Streptococcus* sp. For use as plasma extender, the high-molecular-weight dextran is acid hydrolyzed, fractionated (as to its molecular weight distribution), and purified. The specifications on the dextran for use as a plasma volume extender vary in different countries; the physical, chemical, and biological specifications it must meet are extremely rigid. *See* INDUSTRIAL MICROBIOLOGY; POLYMERIZATION.

Henry M. Tsuchiya

Diabetes

A condition in which excessive amounts of some substances are excreted from the body. The term may refer to either of two unrelated diseases, diabetes mellitus and diabetes insipidus. The word diabetes derives from the Greek for siphon, a reference to the copious urine excretion that characterizes this affliction. In common usage, the term diabetes is synonymous with diabetes mellitus.

Diabetes Mellitus

Diabetes mellitus is a disease of abnormal carbohydrate metabolism in which glucose cannot enter the body's cells and be utilized, and therefore remains in the blood in high concentrations. Mellitus is the Latin word meaning honey, and is applicable because the urine of patients with diabetes mellitus has a sweet taste, distinguishing this disease from diabetes insipidus, in which large volumes of dilute, almost colorless and tasteless urine are passed. In diabetes mellitus the excess sugar in the blood (hyperglycemia) leads to the excretion of sugar in the urine (glycosuria), a cardinal diagnostic symptom. Glycosuria in

turn causes the excretion of large amounts of urine (polyuria), which results in dehydration and intense thirst (polydipsia). Although blood glucose is high, it cannot enter the appetite-regulating cells of the hypothalamus; hunger is therefore great, and the diabetic person tends to eat constantly (polyphagia). But because glucose cannot enter and nourish the cells, body tissues are subjected to the equivalent of starvation; rapid weight loss occurs, part of which is due to the excretion of water in urine. *See* CARBOHYDRATE METABOLISM.

Varieties. Diabetes mellitus appears in two varieties, each with its own cause: diabetes mellitus type I (formerly known as juvenile onset diabetes), caused by deficiency of the pancreatic hormone insulin (whose chief function is to promote the entry of glucose into cells); and diabetes mellitus type II (formerly known as maturity onset diabetes), in which insulin is available but cannot be properly utilized.

Diabetes mellitus type I. Of the two forms of diabetes, type I is initially the more serious and less common, afflicting approximately 1 in every 600 children. The incidence varies among countries. The highest incidence is seen in Finland and Sweden, and some of the lowest in Korea and Mexico. The United States and Canada are included among the highest. Although it usually appears before the age of 20, type I diabetes can strike at any age. It is caused by a significant shortage or complete lack of insulin secretion. The endocrine cells normally responsible for insulin production are found in the pancreas, but have nothing to do with the ordinary digestive functions of that organ. They occur as islands of endocrine tissue, called islets of Langerhans, scattered throughout the substance of the pancreas. The islets consist of at least three cell types, each of which secretes a different hormone important in regulating carbohydrate metabolism. The alpha cells produce glucagon, which elevates blood glucose; the beta cells produce insulin; and the delta cells produce somatostatin, a hormone that appears to participate in controlling the activity of the alpha and beta cells.

In type I diabetes the beta cells are destroyed, possibly by the attack of the body's own immune system or by virus infection; victims of type I require daily insulin injections to survive. Insulin must be injected rather than taken orally because it is inactivated by the enzymes of the digestive system. Without insulin therapy, a patient with type I diabetes will inevitably deteriorate into a condition of ketosis (presence in the blood of ketone bodies, the metabolic intermediates of fat metabolism) and acidosis (dangerously acidic blood pH). Ketoacidosis, as the combined condition is called, leads rapidly to coma and death in the untreated diabetic.

When Frederick G. Banting and Charles H. Best succeeded in purifying insulin from animal pancreatic tissue in 1921, type I diabetics were given their first chance at living a normal life-span. Today type I diabetics are routinely treated by injections of insulin. Human insulin became increasingly available in the 1980s through the use of recombinant deoxyribonucleic acid (DNA) technology. In this technique the gene specifying human insulin is spliced into the chromosome of a bacterium, and the billions of progeny bacteria then produce insulin as though it were one of their own bacterial proteins, thus serving as minifactories for the hormone. The bacterium used to produce insulin is *Escherichia coli*. Another DNA technique for producing insulin uses yeast instead of *E. coli*. Up to the early 1980s, insulin was extracted from cattle and pigs. This "foreign" insulin is still used, but has some unwanted side effects not seen or rarely seen with human insulin, therefore making human insulin the treatment of choice. One of the most important characteristics of insulin produced by genetic engineering, other than its being identical to human pancreatic insulin, is its limitless availability. *See* GENETIC ENGINEERING; INSULIN; PANCREAS.

Diabetes mellitus type II. Type II is the more common form of diabetes mellitus, accounting for 8 out of every 10 cases. It usually appears after the age of 30, although an uncommon variant called type II diabetes of the young occasionally occurs in obese children and teenagers. The initial symptoms of type II diabetes are much less noticeable than those of type I, and the characteristic triad of polyuria, polydipsia, and polyphagia may even be entirely absent. For this reason, type II diabetes can exist undetected for dangerously long periods. Victims are not prone to the ketoacidosis and coma that threaten type I patients, and their disease can usually be managed without insulin injections. Thus two other names for type II diabetes are nonketosis-prone diabetes and non-insulin-dependent diabetes.

Whereas type I diabetes is a disease of insulin shortage, victims of type II diabetes usually have insulin in their bloodstream. In fact, insulin levels in type II diabetics are sometimes even higher than those in nondiabetic individuals. However, since the cells of a type II patient do not respond to insulin by taking in and utilizing blood glucose as normal cells would, the type II diabetic may have hyperglycemia in spite of high insulin levels. The probable reason for this defect is that the cells have specific receptor molecules on their surface membranes whose job is to recognize insulin and trigger the biochemical steps leading to glucose uptake and utilization. Interestingly, the number of insulin receptors on cells can be changed by gaining or losing body weight; the more body fat a person carries, the fewer insulin receptors there will be on the cells. This may explain why type II diabetes is mainly a disease of the obese, and why weight reduction is such effective therapy. High glucose levels also decrease the number of insulin receptors; normalizing glucose levels is thus of utmost importance.

The most important treatment for most cases of type II diabetes consists of exercise and weight reduction, which will return glucose metabolism to normal in 9 out of 10 cases. Because the cells of type II diabetics lack responsiveness to their own insulin, insulin injection therapy is used for some of these cases. Approximately 10% of all type II patients must use insulin. This percentage appears to more than double in Black, Latino, Native American, or other minority patients living in the United States. In

addition, a family of drugs known as oral hypoglycemic agents is often used in treating the disease. These drugs, including tolbutamide, chlorpropamide, acetohexamide, and tolazamide, lower blood glucose by stimulating the pancreatic cells to release insulin, and by augmenting insulin's effect on the cells of the body. All of these drugs are known as first-generation sulfonylureas. However, there are other drugs with the same basic molecule but with chemical changes in their structure; these are called second-generation sulfonylureas. This group of drugs is safe if properly used and is effective in many cases if used in conjunction with proper diet and exercise programs.

Amylin. Amylin is a 37-amino-acid peptide hormone that occurs naturally in humans, and has been shown to be in excess in individuals with non-insulin-dependent diabetes mellitus. Amylin has been shown to be deposited in the islets of Langerhans, and may decrease insulin release and insulin production in the pancreas, thus causing a deterioration in glucose control in non-insulin-dependent diabetes.

In insulin-dependent diabetes, amylin appears to slowly decrease with the progression of the disease, and when none is found, individuals may be more prone to have hypoglycemia unawareness (low blood glucose levels without symptoms).

Complications. Both type I and type II diabetes tend to produce a number of debilitating and life-threatening complications despite the use of insulin or oral hypoglycemic agents. It is primarily these complications, not the risk of ketoacidosis, that account for the heavy burden this disease inflicts on society, and make diabetes the third leading cause of death in the United States.

Much evidence indicates that most of the complications of diabetes are caused by the elevated blood glucose levels common in this disease. Even diabetics who take regular insulin injections exhibit wide fluctuations in their blood glucose levels, in marked contrast to the tightly controlled blood glucose levels observed in nondiabetics. According to the prevalent hypothesis, the more diabetics succeed in controlling their blood glucose to near-normal levels, the less risk there will be of developing any complications of the disease. Thus, careful attention to diet, exercise, regular medication, and monitoring of blood glucose is vitally important for any diabetic.

The most common complications of diabetes are the following.

Eye damage. Diabetes is the leading cause of nontraumatic blindness in young adults. The blood vessels of the diabetic eye are prone to microaneurysms (small outpouching of the vessel wall), excessive vascular permeability (leakage from the vessel), growth of newer, more fragile vessels, and ultimately hemorrhage (bleeding). These events can lead to scarring, fibrosis, and subsequent loss of vision.

Kidney damage. In long-standing diabetes, the smallest blood vessels of nephrons (the functional filtration units of the kidney) frequently suffer the same type of damage as the blood vessels of the retina; this condition is called diabetic nephropathy. Severe diabetic nephropathy commonly leads to kidney failure, and such patients require dialysis in order to survive. The prognosis for individuals with diabetes on dialysis is poor, with only 30% surviving beyond 5 years.

Premature atherosclerosis. Atherosclerosis (often called arteriosclerosis or hardening of the arteries) is a condition in which arteries become progressively occluded by accumulations of cholesterol-containing plaque. Although atherosclerosis is common in older nondiabetics, diabetics tend to develop it earlier and in a more severe form. The most common consequence of atherosclerosis is coronary artery disease and stroke, which result when the blood supply to the heart or the brain is obstructed by a plaque within an artery. Coronary artery disease is the leading cause of death in long-term diabetics, accounting for 75% of fatalities. Also, because of atherosclerosis and restricted blood flow, diabetics are often subject to gangrene, especially of the extremities. *See* ARTERIOSCLEROSIS.

Nervous system abnormalities. Patients with diabetes may experience a variety of nervous system disturbances, collectively referred to as diabetic neuropathy. Symptoms may include muscle weakness, pain in the extremities, local paralysis, urinary incontinence, sexual impotence, and sensations of cold, heat, or tingling in various parts of the body. There may also be episodes of diarrhea, frequently nocturnal, which is a characteristic of the "diabetic diarrhea" problems involving the upper intestinal tract. Nervous system involvement of the esophagus and stomach may cause nausea and even vomiting.

Causes. Three factors that are believed to be important in the causation of this disease are heredity, viral infections, and immunological damage to the pancreas.

Heredity. Although diabetes is popularly considered a genetic disease, it is not inherited in the sense that one can inherit curly hair. Heredity does operate in determining one's risk of diabetes, but only as a predisposing factor, not as an absolute determinant. The two types of diabetes mellitus have entirely different patterns of inheritance, and the presence of one type in a family does not influence one's risk of developing the other. Genetic factors play a much more important role in type II diabetes than in type I diabetes. For example, the identical twin of a person who develops type II diabetes after age 40 is almost 100% certain to develop the disease, but someone whose identical twin has type I diabetes runs only a 52% risk of becoming a diabetic. Parents with type I diabetes face only a 2% risk of having a child with the disease, whereas the risk to the offspring of a type II diabetic is about 10%. Genetic predisposition is apparently stronger in the case of type II diabetes of the young, with a 50% chance that the child of such a diabetic will also develop the disease. The environment obviously plays a significant role in type II diabetes. *See* HUMAN GENETICS.

Virus infections. Virus infection is strongly implicated as one causative factor in the beta-cell destruction that characterizes type I diabetes, but it is not involved in type II diabetes. Some of the viruses that have been implicated are the agents of mumps and

rubella (German measles), whose action seems to be delayed for months or years, and coxsackievirus B, which is more likely to precipitate diabetes shortly after the person recovers from the influenzalike infection. Apparently, several factors operate to determine whether or not diabetes will develop in a person infected with one of these extremely common viruses. They include the organ affinity of the particular virus strain, the person's susceptibility to viral invasion of the pancreatic tissue, and the individual's capacity to regenerate islet tissue that has been damaged by a viral infection. *See* ANIMAL VIRUS.

Immunological damage to pancreas. Autoimmune reactions, wherein the body's immune defense system attacks its own pancreatic tissue as though it were a foreign substance, are also suggested as a cause for the beta-cell destruction of type I diabetes. An autoimmune response may be conducted by T lymphocytes or by soluble antibodies in the blood or by both. These reactions may arise spontaneously or may be secondary to a viral infection of the pancreas that leaves the islet tissue modified in some way that marks it for attack. *See* AUTOIMMUNITY.

Diet. Dietary management has been part of diabetic therapy since preinsulin days, when starvation diets were used to prolong the life of type I diabetics. More recently, weight reduction for victims of type II diabetes and complex carbohydrate liberalization for those of both types of diabetes mellitus have been the cornerstone of therapy.

For obese type II diabetics, weight reduction is certainly appropriate, since it tends to increase the number of insulin receptors on the cells of the body, usually making it possible to normalize blood glucose without resorting to the use of insulin injections or oral drugs.

Some other dietary measures besides reducing caloric intake are now believed beneficial for type II diabetics. The ingestion of foods high in fiber content, for example, has been shown to reduce hyperglycemia in type II diabetics.

For type I diabetics, the traditional diet is high in protein and quite restricted in carbohydrates, with only about 45% of the total daily caloric intake being carbohydrate-derived. Since most type I diabetics are thin, weight reduction is not part of their program. Presently, the recommended carbohydrate intake has been liberalized to 55–60% of the diet, plus 15–20% protein and 20–30% fat. There has been a trend toward relaxing the carbohydrate restriction for type I diabetics as evidence accumulates that diets containing even as much as 80% carbohydrates can be compatible with excellent control of diabetes, as long as the carbohydrates are complex (that is, whole grains, legumes, potatoes, bread) rather than refined (that is, sugars, candy, soft drinks, desserts, and so on). Artificially sweetened food can be very beneficial for weight control; however, it is not a necessary component of treatment.

Prospects. The ideal solution is to be able to mimic insulin secretion, with intensive insulin therapy used with either multiple insulin injections or insulin infusion pump therapy. The insulin pump continuously delivers a small amount of insulin, and the individual can increase the amount before meals based on his or her glucose level. Still, this setup is not ideal: in a nondiabetic, insulin is delivered directly into the portal blood system, while therapy delivers insulin under the skin, where it can be absorbed into the bloodstream. Hooking a glucose-monitoring device to the insulin pump would allow the system to close, permitting the individual to have an artificial pancreas implanted (several models are available).

Healthy pancreatic tissue can be transplanted from a donor into a type I diabetic, but there are many problems related to this method, such as the typing of the pancreatic tissue from the donor, the possibility of rejection, and the possibility of infections.

Diabetes Insipidus

Diabetes insipidus is caused by a deficiency of, or resistance to the action of, vasopressin, the antidiuretic hormone produced by the posterior lobe of the pituitary gland. If the pituitary fails to produce vasopressin, the condition is called central diabetes insipidus. If the kidneys do not respond to the vasopressin and fail to concentrate urine, the condition is labeled nephrogenic diabetes insipidus. Symptoms include increased thirst and frequent urination. Permanent damage to the kidneys can result if the condition is not treated.

A low level of vasopressin can be corrected with desmopressin acetate, a synthetic analog of 8-argi-ninevasopressin, which can be administered in the form of a nasal spray. Desmopressin acetate is ineffective in the treatment of nephrogenic diabetes insipidus. *See* NEUROHYPOPHYSIS HORMONE.

Jaime Davidson; Peter E. Bressler

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Diagenesis

All the chemical, biochemical, and physical changes that sediments undergo from the time of deposition until the stage of metamorphism is reached. Diagenetic changes are gradational with metamorphism at elevated temperatures or pressures, and with atmospheric weathering effects when sedimentary rocks are exposed on the Earth's surface. Diagenesis commonly transforms sediments into sedimentary rocks and results in significant chemical and mineralogical changes. With increased burial, sands undergo grain rearrangement and compaction, mineral dissolution or replacement, and mineral precipitation as cements. Clays formed during weathering processes are transformed into illite, chlorite, and muscovite. Porous carbonate sediments consisting of unstable aragonite or magnesium-calcite are transformed to nonporous limestones composed of calcite. In spite

of these profound changes, many features such as macroscopic sedimentary structures may be preserved in both detrital (sandstones and shale) and carbonate rocks. Diagenetic processes and products vary with original sediment composition and depositional environment, depth of burial (pressure and temperature), burial time, nature and movement of interstitial fluids, and physical, chemical, and biochemical processes.

While all of the variables mentioned above are important in controlling diagenesis, the one that received the most attention in the 1990s was interstitial fluid flow. At shallow subsurface depths, the sediment-water system is open, and diagenetic processes are driven by the chemistry of pore waters that migrate through the sedimentary units. The majority of these waters are meteoric or derived from compacting shale units. At greater depths, permeabilities are reduced, and the system becomes progressively more closed. Under these conditions, the composition of framework grains may be the dominant control on diagenesis. The system can also be reopened at depth as a result of fluids introduced by clay transformations, by hydrocarbon maturation, or along fault and fracture systems (**Fig. 1**).

Processes. Diagenetic processes include purely physical ones that involve rearrangement of the sediments, such as compaction, slumping, bioturbation by organisms, infiltration, and soft sediment deformation; biochemical or organic processes such as particle accretion, boring, and decomposition; and physiochemical processes such as cementation, authigenesis (formation of new minerals), inversion, recrystallization, grain growth, replacement, and dissolution.

Diagenetic processes are also affected by microbial activity, primarily of aerobic and anaerobic bacteria, that can influence the pH and the oxidation potential (Eh) of sediment pore water and the accumulation of chemical reactants. Anaerobic bacteria may survive in sedimentary basins to depths of several thousands of meters below the Earth's surface. Sulfate-reducing bacteria, for example, may interact with the sulfate radicals in dissolved formation waters, producing hydrogen sulfide, calcite, native sulfur, and pyrite, among other products important in diagenetic reactions.

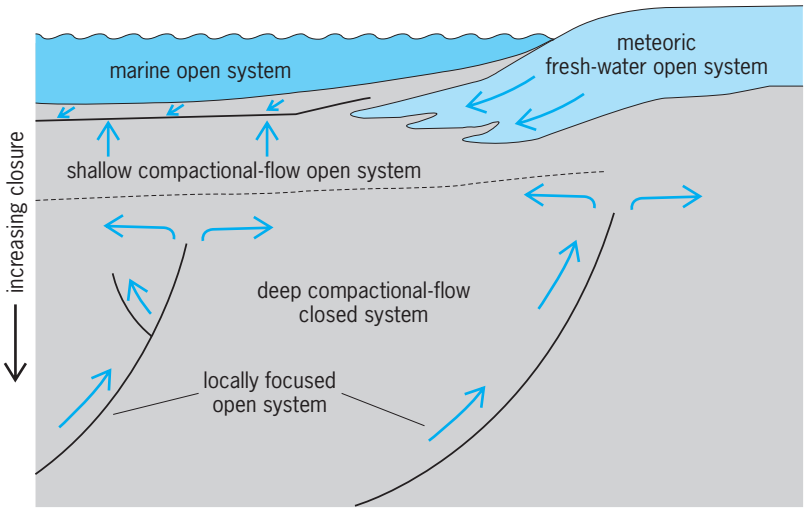

Fig. 1. Subsurface flow systems. Diagenetic processes occur in open systems at shallow subsurface depths (meteoric fresh-water open systems and shallow compactional-flow open systems). At greater depths, decreasing flow rates promote closure below approximately 1 km (3300 ft) [deep compactional-flow closed systems]. Deep-burial open-system conditions are restricted to areas associated with fault or fracture systems. (*From M. D. Wilson, ed., Reservoir Quality Assessment and Prediction in Clastic Rocks, SEPM Short Course, no. 30, 1994*)

Most diagenetic processes involve a reduction in porosity and permeability. An understanding of these two sediment properties is critical to understanding the migration of subsurface fluids and the accumulation of oil and gas and certain types of mineral deposits in subsurface rock units. However, one diagenetic process, dissolution, is of major importance in creating secondary porosity. Only a few of the more important processes will be reviewed here.

Although compaction and cementation probably constitute the principal processes responsible for loss of porosity in sediments, several processes that operate penecontemporaneously with deposition may lead to loss of porosity, even before compaction and cementation begin. These processes include bioturbation and mechanical infiltration. Bioturbation is the disturbance of the original sediment fabric by the activity of animals and plants. The principal activity is burrowing by invertebrate or small vertebrate organisms. Burrowing usually results in a worsening of sorting from the mixing of sand and clay layers and a corresponding reduction in porosity in the sand layers.

Mechanical infiltration involves the downward movement of clay-size particles carried by percolating meteoric waters. This process is particularly important in coarse-grained desert deposits where water tables are low and streams are influent during rare periods of runoff. Infiltrated clays are recognized by their clastic texture and the fact that they tend to accumulate on the upper surface of grains in geopetal fabric.

As sediments are buried, the weight of overburden or tectonic stress results in compaction. During compaction, grains move closer together, pore space is reduced, and pore waters are expelled. The amount of compaction achieved is a function of the nature of the sediment, the depth of burial, and the degree of cementation. Compaction of sediments is usually achieved by one or more of the following processes: rotation and slippage of grains; plastic or ductile, flexible, and brittle grain deformation; and pressure solution (**Figs. 2** and **3**).

Cementation involves the precipitation of authigenic minerals from aqueous solutions in the pore spaces of sediment. Geochemical calculations suggest that large volumes of pore fluids must pass through the rock to account for the observed authigenic minerals. The most commonly documented authigenic cements are carbonates, silica (quartz), clays, and zeolites. *See* AUTHIGENIC MINERALS.

Three closely related processes important in the diagenesis of limestone are inversion, recrystallization, and grain growth. Inversion is the process by

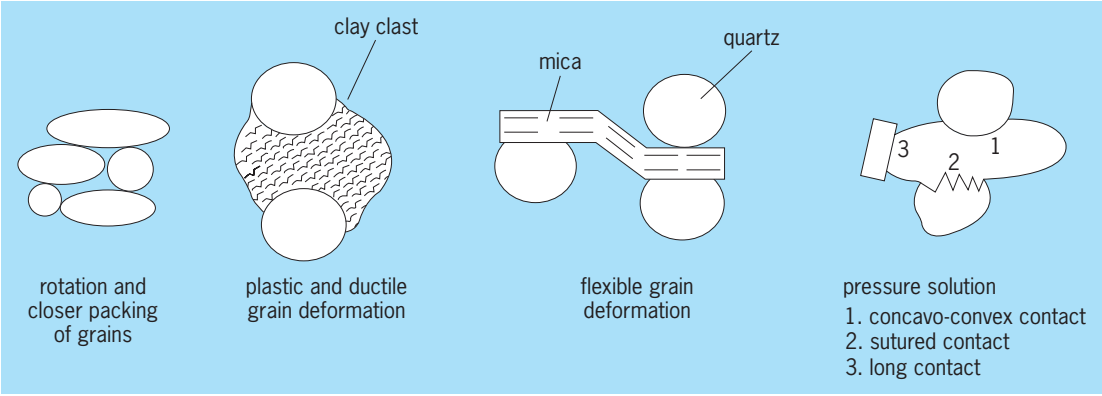

Fig. 2. Some methods of porosity loss resulting from compaction, including rotation and closer packing of grains; plastic and ductile grain deformation; flexible grain deformation; and pressure solution. (*From M. D. Wilson, ed., Reservoir Quality Assessment and Prediction in Clastic Rocks, SEPM Short Course, no. 30, 1994; and J. C. Wilson and E. F. McBride, Compaction and porosity of Pliocene sandstones, Ventura basin, California, Amer. Ass. Petrol. Geol. Bull., 72:663–681, 1988*)

which unstable minerals are converted to a more stable form with no change in chemical composition but with a different crystal structure. The most commonly documented example is the inversion of aragonite to calcite. Recrystallization and grain growth refer to processes that result in a change in crystal or grain size without chemical alteration. The distinction of which process was responsible for the resulting texture seen in many ancient limestones is not always possible. Neomorphism includes all three processes. *See* ARAGONITE; CALCITE.

Interstratal solution involves the dissolution of mineral grains and cements (Fig. 3) from strata by migrating pore fluids or ionic diffusion. This process is locally important in the formation of many oil and gas reservoirs in both limestones and sandstones. Clues to the recognition of secondary porosity in sandstones include oversized pores, partially dissolved and corroded grains, enlarged pore throats, and floating grains.

Sandstones. Three major groups of sandstones have been recognized: quartz arenites, arkosic arenites, and lithic arenites. Most quartz-rich sandstones (quartz arenites) are partially or completely cemented by silica and calcite. Secondary silica commonly occurs as overgrowths in optical continuity with original grains. Considerable controversy remains over the source of authigenic silica. Possible sources include pressure solution, illitization of smectitic clays, siliceous microfossils, dissolution of volcanic glass, feldspar alteration and dissolution, and carbonate replacement of silicate minerals. Figure 3 suggests that quartz overgrowths occur at relative low rates and over a very wide range of depths.

Carbonate cements, including calcite, dolomite, siderite, and ankerite, are the most common and diverse cement in sandstones. Several different generations of carbonate cement can be found in one sandstone formation, and they may display a considerable range of crystal forms, sizes, and distributions. The abundance of calcite cement results, in part, from the fact that connate water in shallow marine sediments is oversaturated with respect to calcium carbonate. Other sources of carbonate ions might be dissolution of fossils, clay membrane filtration, and pressure solution of interbedded limestones. *See* SANDSTONE; SILICATE PHASE EQUILIBRIA.

Clay minerals and zeolite cements may be abundant in mineralogically immature arkosic and lithic sandstones that are rich in feldspar and volcanic or other highly unstable rock detritus. These authigenic clays have delicate, crystalline morphologies and occur as rim cements on grains, as occluding cements between grains, and as fracture or vug fillings. Common authigenic clay minerals include

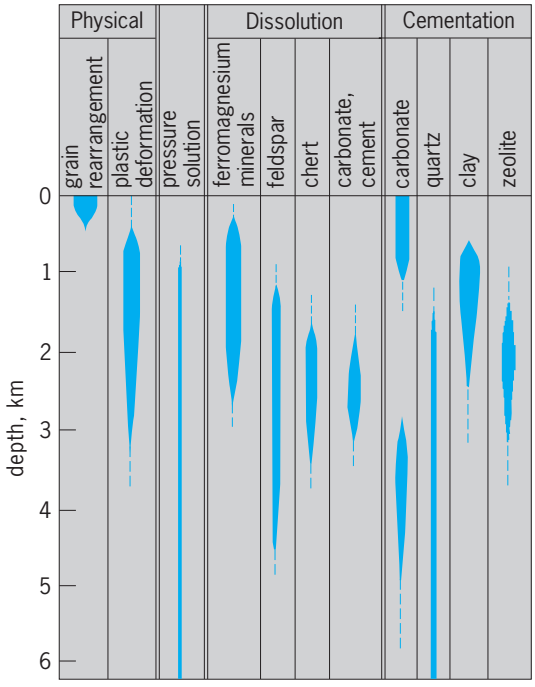

Fig. 3. General depth regimes for major diagenetic processes in detrital sediments and sedimentary rocks. Physical processes tend to dominate at shallow depths, and chemical processes at greater depths. Width of bars indicates the relative rate at which processes operate. (*From M. D. Wilson, ed., Reservoir Quality Assessment and Prediction in Clastic Rocks, SEPM Short Course, no. 30, 1994*)

chlorite, illite, smectite, kaolinite-dickite, and mixed-layer illite-smectite. Zeolite pore-fill cements may accompany these diagenetic clays in volcanic sandstones. The principal control on the development of diagenetic clays and zeolites in mineralogically immature sandstones is probably temperature, which is controlled by depth of burial. Not all immature sandstones contain abundant authigenic clay minerals or zeolites. Some sandstones contain lithic fragments, which deform plastically to produce a pseudomatrix, thus reducing permeability, restricting fluid movement, and retarding development of authigenic clay minerals and zeolites.

Three other diagenetic processes are important in felspathic sandstones. In fine-grained feldspathic sandstones, with little clay matrix, diagenetic potassium feldspars form as overgrowths on feldspar grains and as euhedral to subhedral crystals filling pore spaces. Potassium- and calcic-rich feldspathic sandstones can be altered by albitization (formation of authigenic sodium-rich albite). The principal reason that albite forms is that most pore waters are rich in sodium ions. The last process of feldspar alteration is calcification, which involves the replacement of feldspars by calcite.

The red color in some arkosic or lithic sandstones is provided by hematite, which is of diagenetic origin. The sediments are not red when deposited, but redden with time. Iron-bearing minerals in these immature sandstones alter by hydrolysis and release iron, which precipitates as hematite or as a precursor oxide which converts to hematite upon aging. *See* HEMATITE; REDBEDS.

Secondary porosity, caused by dissolution of primary or secondary minerals, is found in sandstones of all ages and depths. In spite of this fact, there remains considerable controversy over the origin, importance, and effect of secondary porosity on hydrocarbon-reservoir and ground-water-reservoir quality. Meteoric water influx is the most important leaching mechanism, and porosity produced by this mechanism may persist during deeper burial, provided that the sandstones do not have high percentages of ductile grains. Mechanisms for the production of secondary porosity at depth include silicate hydrolysis and organic acids. The quantitative importance of these various mechanisms is unclear.

Shales. Diagenetic changes in shales involve color changes, compaction, decomposition of unstable minerals and recrystallization of others, and development of micas of restricted composition. Clay-mineral components of fine-grained sediments show evidence of diagenesis, starting with ionic substitution in clay mineral lattices, and ending with complete conversion from one clay mineral type to another, as in the case of the conversion of smectite to illite. One theory that explains the mineral alteration of smectite to illite is known as the crystal growth theory, and suggests that this conversion not only involves the addition of potassium to the smectite crystals but also requires that the crystals grow in size.

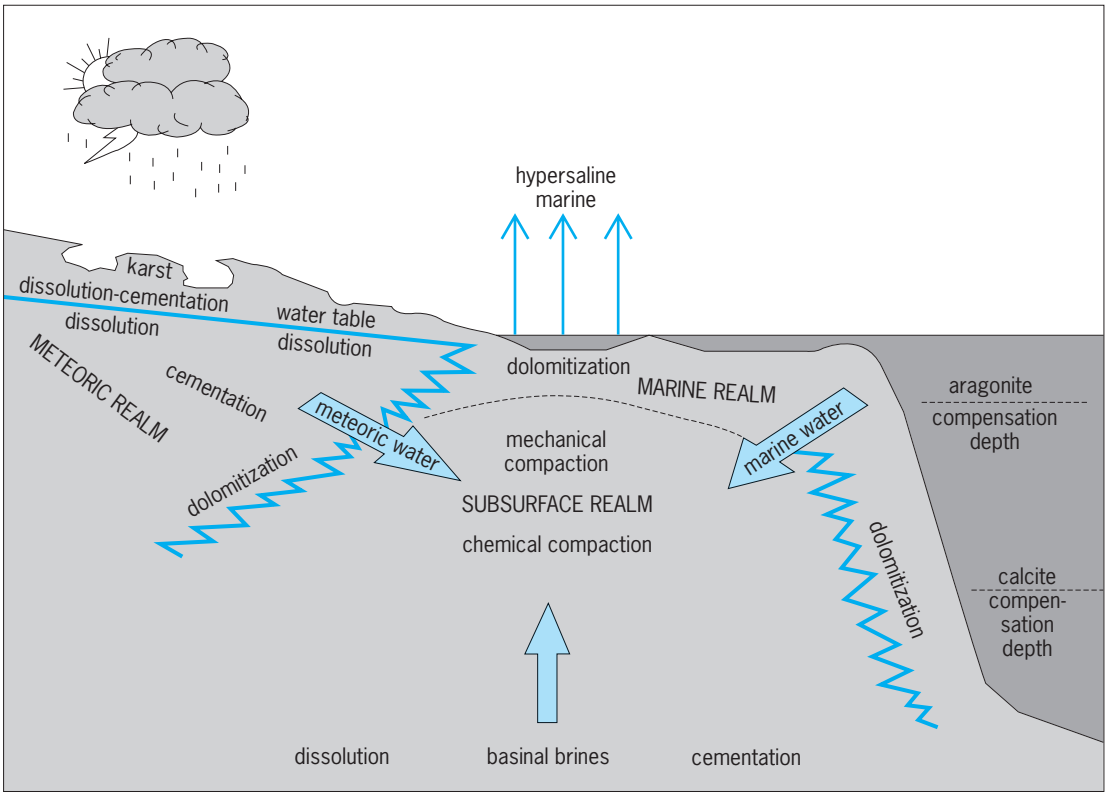

Fig. 4. Principal subsurface realms in which diagenesis of carbonate sediments occur. Major diagenetic processes that operate in each of the realms (meteoric, marine, and subsurface) are indicated. (*From C. H. Moore, Carbonate Diagenesis and Porosity, Elsevier, Amsterdam, 1989*)

Compaction is another important process in shales. Diagenetic changes accompanying compaction and mineral alteration with increased depth of burial and corresponding increases in pressure, temperature, and time result in a reduction in porosity and permeability, and the expulsion of both water and adsorbed hydrocarbons from the shales. The processes are therefore considered important to the origin of petroleum. Sediments with high percentages of clays usually contain large quantities of water at the time of deposition. Pore waters are usually expelled first during the early stage of compaction as platy clay particles pack together to form a less open framework. During the initial stage, original sediment volume is reduced by 50% or more. After this degree of compaction is achieved, further compaction is hindered by the small size of the pore diameters, and by the fact that most of the remaining waters are attached to the clay surfaces. Final expulsion of the remaining waters and any accompanying hydrocarbons may be accomplished only after the final mineral alteration of smectite to illite at burial between 6000 and 8000 ft (1800 and 2400 m). In some thick shale sequences, however, compaction may be delayed because an impermeable seal prevents the escape of pore waters. In these sequences, abnormally high pressures may build up. *See* CLAY MINERALS; ILLITE; SHALE.

Carbonates. Carbonate sediments are recognized as being especially susceptible to diagenetic modification. Carbonate diagenesis occurs in three regions or realms (meteoric, marine, and subsurface; **Fig. 4**). Marine carbonate sediments can be brought into contact with meteoric water (fresh water) by a sea-level fall or tectonic uplift. Meteoric waters are highly aggressive and readily dissolve aragonite and high-magnesium calcite. Dissolution of these two minerals may saturate pore waters with respect to calcium carbonate, resulting in precipitation of calcite cement. Karst topography is commonly generated on subaerially exposed marine carbonates by dissolution. *See* KARST TOPOGRAPHY.

On the sea floor and in the shallow-marine subsurface (the marine real), the principal diagenetic processes are bioturbation, modification of shells by boring organisms, dissolution and cementation of grains, and penecontemporaneous dolomitization. This last process is characteristic of nearshore, hypersaline environments, and takes place while the host sediment is in its original depositional setting (Fig. 4).

As carbonate sediments and rocks are buried, pressure and temperature increases and the pore fluids change. Eventually the carbonates reach the subsurface realm and undergo physical and chemical compaction and additional dissolution, cementation (Fig. 4), and replacement as dictated by the specific conditions (pore-fluid composition, temperature, pH, and Eh) of the subsurface realm. Dolomite rock (dolostone) is a common and important replacement product that occurs in this realm. Beneath modern carbonate islands, this type of dolomite is more common and pervasive than penecontemporaneous dolomites, and may provide an analog for the extensive dolostones found in the geologic record. Important hydrocarbon reservoirs are found in many of these dolostones where replacement of calcite by dolomite has resulted in the formation of intercrystalline porosity. *See* CARBONATE MINERALS; DEPOSITIONAL SYSTEMS AND ENVIRONMENTS; DOLOMITE ROCK; SEDIMENTARY ROCKS; SEDIMENTOLOGY.

Frank G. Ethridge

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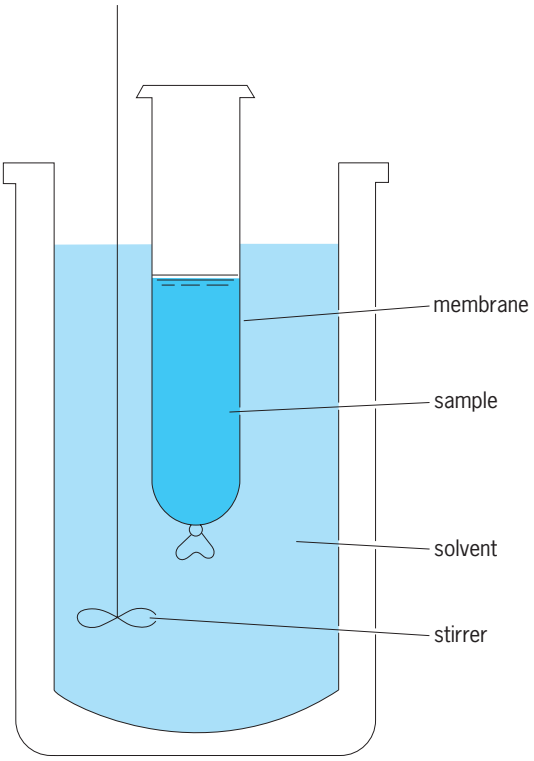
Dialysis

A process of selective diffusion through a membrane by dissolved solutes in liquid solution. As dialysis is usually carried out, the membrane permits the diffusion of low-molecular-weight solutes (crystalloids) but prevents the passage of colloidal and high-molecular-weight solutes (macromolecules). Membranes suitable for this purpose include vegetable parchment, animal parchment, goldbeater's skin (peritoneal membranes of cattle), fish bladders, dialyzing cellophane (Visking sausage casing), and collodion (nitrocellulose deposited from alcohol-ether solution).

The solution is contained within such a membrane. The low-molecular-weight solutes are removed by placing pure solvent outside the membrane. This solvent is changed periodically or continuously until the concentration of diffusible solutes in the solution is reduced to near zero. The technique is used extensively in separating and purifying macromolecules of biological origin (see **illus.**).

Dialysis rates of ionic low-molecular-weight solutes may be increased greatly by applying an electric field to achieve an electrophoretic movement of ions through the dialytic membrane. This combined process of dialysis and electrophoretic transport of solutes through a membrane is known as electrodialysis. *See* ELECTROPHORESIS.



Equipment for dialysis.

Ions do not migrate readily through membranes that carry in their pores charges of the same sign as the ion; ions of charge opposite to that of the membrane are not prevented from passing through. Such a selective permeability to ions by charged membranes has been used to improve the efficiency of electrodialysis. Membranes of collodion, vegetable parchment, and cellophane carry negative charges when in contact with aqueous solutions. Animal membranes show positive charges at low pH and negative charges at high pH.

Membranes of ion-exchange resins are commercially available in cationic and anionic forms. The ion-exchange membranes show a high selectivity for ions of one charge type. *See* ION-SELECTIVE MEMBRANES AND ELECTRODES.

For desalting small quantities of solutions for chromatography and radioisotope tracer studies, micro- and semimicroelectrodialyzers have proved valuable. On a larger scale, electrodialysis combined with ion-exchange membranes has been used to remove salt from seawater. *See* COLLOID; WATER DESALINATION.

Quentin Van Winkle

Diamagnetism

That branch of magnetism which treats of diamagnetic phenomena and of the properties of diamagnetic bodies. Diamagnetism is a property exhibited by substances with a negative magnetic susceptibility, that is, by substances which magnetize in a direction opposite to that of an applied magnetic field. A diamagnetic substance has a magnetic permeability less than 1, and is repelled when placed near a magnet. The magnetization of diamagnetic substances is associated with the currents induced on application of a magnetic field. According to Lenz's law, the flow of an induced current is in such a direction as to oppose the change of flux of the inducing field; this accounts for the negative susceptibility. The diamagnetic susceptibility is invariably small, of the order of -10^{-5} cm^3/mole. *See* LENZ'S LAW; MAGNETIC SUSCEPTIBILITY.

All matter responds to applied fields in this diamagnetic fashion. However, some substances also have net electronic orbital or spin magnetic moments, or both, which can be aligned by an applied magnetic field in direction along (not opposite to) the field; this property is called paramagnetism. For these substances, as shown in Eq. (1), the observed susceptibility χ is the sum of diamagnetic and paramagnetic terms. Under ordinary conditions χ_d is temperature-independent. Hence, if χ_p follows the inverse temperature dependence of Curie's law, one can experimentally determine the separate contributions of χ_d and χ_p by measuring χ as a function of temperature. However, if χ_p is also temperature-independent, as in metals and in Van Vleck paramagnetism, the problem of determining the separate contributions becomes quite difficult. *See* PARAMAGNETISM.

$$\chi = \chi_d + \chi_p \tag{1}$$

Although all matter exhibits diamagnetism, only those substances in which $\chi_p = 0$ are referred to as diamagnetic. This is because paramagnetism, if present, usually predominates, and the gross magnetic response of the material is paramagnetic. Important exceptions are the alkali and alkaline earth metals, where χ_p is unusually small and of the order of χ_d, and salts or solutions containing only a small fraction of paramagnetic atoms. The condition for $\chi_p = 0$, and hence pure diamagnetism, is that all electronic spins be paired and all orbital moments either be zero or effectively cancel one another. Nearly all molecules with an even number of electrons satisfy this condition; an important exception is O_2. The condition is also satisfied by most nonmetallic solids, except compounds containing atoms with incomplete inner-shell electron groups, such as the transition, rare-earth, and actinide elements. *See* ELECTRON SPIN.

As stated previously, the diamagnetic response of a substance is small; only a very small fraction of the applied magnetic field is shielded from the interior of the substance by the induced diamagnetic currents. There is one case, however, in which the inducing field is completely shielded (except for small surface effects). This is the perfect diamagnetism exhibited by superconductors, and is known as the Meissner effect. *See* MEISSNER EFFECT; SUPERCONDUCTIVITY.

Langevin theory. The Langevin theory of diamagnetism is based on the idea of inducing an electronic current inside an atom. The theory employs the Larmor theorem. *See* LARMOR PRECESSION.

In a magnetic field H, the precession of the Z electrons within the atom is equivalent to a current equal to $-Z(e/c)(\omega_L/2\pi)$ in electromagnetic units. Here e/c is the magnitude of the electronic charge in emu, and ω_L is the angular Larmor frequency as shown in Eq. (2), where m is the electronic mass. The mag-

$$\omega_L = \frac{-eH}{2mc} \tag{2}$$

netic moment μ arising from this induced current is equal to the product of the current and the area of the current loop, as in Eq. (3), where $\overline{\rho^2}$ is the sta-

$$\mu = \frac{-Ze}{c}\frac{\omega_L}{2\pi}\overline{\rho^2} \tag{3}$$

tistical average, over a large number of atoms, of the square of the perpendicular distance of an electron from the field axis. This average is equivalent to $\overline{x^2} + \overline{y^2}$ if H is along z. For a random assembly of atoms, since $\overline{x^2} = \overline{y^2} = \overline{z^2}$, one may write Eq. (4), where $\overline{r^2}$ is the mean-square distance of the electron from the

$$\overline{\rho^2} = 2/3\left(\overline{x^2} + \overline{y^2} + \overline{z^2}\right) = 2/3\,\overline{r^2} \tag{4}$$

nucleus. Thus, the diamagnetic susceptibility of N atoms is given by Eq. (5). This is P. Langevin's result,

$$\chi_d = \frac{N\mu}{H} = -\frac{Ze^2N}{6mc^2}\,\overline{r^2} \tag{5}$$

as corrected by W. Pauli. The molar susceptibility χ_M is obtained by replacing N in Eq. (5) by Avogadro's number. Numerically, it may be expressed as Eq. (6),

$$\chi_M = (-2.83 \times 10^{10}\ \text{cm/mole})\ \Sigma\overline{r^2} \tag{6}$$

where the summation is to be taken over all the electron orbits in the atom. Since $\overline{r^2}$ is of the order of 10^{-15} cm^2, this gives χ_M of about -10^{-5} cm^3/mole.

Langevin's formula is not modified by quantum mechanics, and the problem becomes that of determining $\overline{r^2}$ of the electronic wave function. The calculation for many-electron atoms is quite involved, and experimental values of $\overline{r^2}$ as determined by using Eq. (5) or (6) give a very useful check of the nature of the wave function for large r. This complements x-ray and electron diffraction data which give information, for the most part, only for small r.

Ionic crystals. In the case of diamagnetism in ionic crystals, the Larmor theorem holds for the individual ions. The diamagnetic susceptibility may be computed with reasonable accuracy from the sum of the individual ion susceptibilities.

Diamagnetism in rare gases and in rare-gas configurations of ions in ionic crystals is shown in the **table**.

Molar diamagnetic susceptibilities (all $\times$ 10^{-6} cm^3/mole)

He −2.0	Li −.07	Mg −4.3	F −9.4
Ne −7.0	Na −6.1	Ca −10.7	Cl −24.2
Ar −19.2	K −14.6	Sr −18.0	Br −34.5
Kr −28	Rb −22.0	Ba −29.0	I −50.6
Xe −43	Cs −35.0		

Molecules. In most molecules the electrons are not moving in a single field of force, and the Larmor theorem breaks down. The total susceptibility of H_2 is given by the sum of Eq. (5) and the Van Vleck paramagnetism, the latter corresponding to the presence of a mean-square magnetic moment, although the mean moment vanishes. Calculations for more complicated molecules are exceedingly difficult. In aromatic ring molecules, such as benzene, with H normal to the ring, the electrons can precess around the ring, or at least in partial ringlike orbits about many nuclei. This gives rise to a much larger susceptibility than is possible when H is parallel to the ring. Crystals with layerlike structures, such as antimony, bismuth, and graphite, also exhibit large anisotropies in diamagnetic susceptibilities.

Bohr–van Leeuwen theorem. This theorem proves the complete absence of magnetism in classical theory. For bound electrons this comes from a cancellation of paramagnetic and diamagnetic susceptibilities, providing one does not (as did Langevin) implicitly quantize the paramagnetic moments by setting them all equal to μ. For a system of free electrons confined to a box, the induced diamagnetic currents in the interior of the box are just canceled by the currents from electrons which bounce in cuspoidal paths off the walls. Thus, magnetism is inexplicable in classical physics and is a quantum phenomenon.

Free electrons. The diamagnetism of free electrons, which vanishes classically, was calculated quantum-mechanically by L. Landau. For particles obeying Fermi-Dirac statistics, such as the electrons in a metal, the numerical value of the Landau diamagnetism is exactly one-third of the Pauli spin paramagnetism.

Bound electrons. The diamagnetism of bound electrons in other than ionic crystals is difficult to calculate. In metals the diamagnetism is a sum of contributions from the nonconducting core electrons, for which the Langevin theory is adequate, and from the conduction electrons, for which the Landau theory must be modified to take account of the periodic potential from the ion cores. Metals in the bismuth group have unusually high diamagnetic susceptibilities (about -10^{-4} cm^3/mole) coming from conduction electrons. These metals also show the de Haas–van Alphen effect, a quasiperiodic variation of susceptibility when plotted aginst $1/H$ at low temperatures. The susceptibility may even oscillate between diamagnetism and paramagnetism.

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Diamond

A mineral composed entirely of carbon; the hardest substance known. Diamond is a polymorph of carbon; lonsdaleite, another polymorph, is sometimes referred to as hexagonal diamond. Diamond is found

on all continents except Antarctica, which has not yet been explored for it. It occurs in nature as single crystals of gem or industrial quality, and as polycrystalline masses referred to as boart, framesite, or carbonado. It has also been found as minute black grains in some meteorites. Diamond can be synthesized in the laboratory and is produced commercially in large amounts for industrial uses. *See* CARBON; GRAPHITE.

Physical and chemical properties. Diamond has a cubic (isometric) crystal structure in which all carbon atoms have covalent (sp^3) bonds. It is this strong bonding that makes diamond hard. Nevertheless, if diamond is struck in specific directions it will readily cleave—a property utilized in the preparation of polished gem diamonds. The combination of refractive index and dispersion gives diamond its brilliance and so-called fire when cut and polished. The thermal conductivity of diamond is the highest of any material (five times that of copper). This property, plus hardness, makes diamond an ideal material for use as a cutting tool in industry and also as a heat sink in electronics. *See* CRYSTAL STRUCTURE.

Although diamond consists of carbon, at least 58 other elements have been found (for example, aluminum, 10 parts per million; hydrogen, 1000 ppm; silicon, 80 ppm) as impurities in natural diamond. However, only two, nitrogen and boron, replace carbon atoms in the diamond lattice. Nitrogen is the major impurity and may substitute for carbon in a number of ways, commonly as either isolated or paired nitrogen atoms, and as discrete platelets of nitrogen within the diamond structure. The presence or absence of nitrogen and the manner of its substitution leads to different physical properties, such as thermal conductivity, electrical resistivity, and infrared spectra (see **table**). Consequently, diamond has been classified into four main types:

Type Ia diamonds contain appreciable nitrogen, often as planar defects (plates) in the crystal structure. They exhibit strong infrared absorption at 1282 cm^{-1} and no paramagnetism. Probably 98% of all natural diamonds are type Ia.

Type Ib diamonds contain less nitrogen than type Ia, and nitrogen occurs as substitutional atoms. Type Ib diamonds are paramagnetic. While type Ib diamonds are rare in nature, most synthetic diamonds are of this type.

Type IIa diamonds contain virtually no detectable nitrogen. These exhibit the highest conductivity of all diamond types. They are very rare in nature.

Type IIb diamonds contain no nitrogen, but boron is present. They are semiconducting, extremely rare in nature, and sometimes blue in color.

Diamond is resistant to chemical attack, other than by strong oxidizing agents. In vacuum or an inert atmosphere, a clear, colorless gem diamond transforms to a gray-black mass of graphite at about 1500°C (2700°F). In air, diamond oxidizes (burns) to carbon dioxide at and above 800°C (1500°F). At high temperature, some metals (for example, tungsten, titanium, and tantalum) react with diamond to form metal carbides. Metals, such as iron, nickel, cobalt, and platinum, in the molten state are solvents for carbon and dissolve diamond; this phenomenon is used as a basis for the synthesis of diamond.

Genesis and occurrence. Most natural diamond, apart from that in meteorites, crystallizes at depths of approximately 110 mi (180 km) in the Earth's upper mantle at temperatures in the range 900–1200°C (1650–2200°F). The host rock in which diamond forms is either a magnesian-rich silica-deficient ultramafic (peridotitic) rock or an ultrabasic eclogitic rock. Minerals that constitute the ultramafic type of diamond-host rock are magnesian rich and include varieties of olivine, pyroxene, and pyrope garnet. The eclogitic rock consists of sodium-bearing pyroxene and an almandine garnet. These various constituent minerals may also occur as inclusions in diamond and result in the host being identified as either an ultramafic (peridotitic) or eclogitic diamond. Diamonds in each group have formed in a distinct and different geochemical environment in the upper mantle. *See* ECLOGITE; LITHOSPHERE; METEORITE; PERIDOTITE.

Dating. Some inclusions in diamond can be characterized as having crystallized simultaneously with the host, and are said to be syngenetic with the diamond. Isotopic analysis of such inclusions enables the age of crystallization of the inclusion, and hence host diamond, to be determined. Some diamonds formed 3.2 billion years ago (the Earth is 4.6 billion years in age), whereas the youngest diamond studied by this method has been dated as 628 million years old.

Diamond is eventually transported to the Earth's surface by unique types of volcanic eruption in which gases play a major role. The eruptions drill narrow (much less than 3000 ft or 1000 m) explosive vents or pipes through the crust of the Earth. Two different rock types, each containing diamond, may result and infill the volcanic neck or pipe. The first type is known as kimberlite, and the second as lamproite. The most productive mine in the world based on the number of diamonds produced per unit of host rock is based on lamproite—the Argyle mine in Western Australia. In general, diamonds are considerably older than the volcanic eruption that transported them to the surface. Thus diamonds that are 3.2 billion years old reached the surface

Selected physical properties of diamond

Property	Value
Crystal structure	Cubic (isometric)
Space group	Fd3m
Unit cell size	$a = 3.57$ A (0.357 nm)
Density	3.52 g/cm^3(2.03 oz/$in.^3$)
Hardness	10 (Mohs scale)
Refractive index	2.417 at 589 nm
Thermal expansion	0.8×10^{-6} at 20°C (68°F)
Thermal conductivity (20°C or 68°F)	
Type Ia	800 W · m^{-1} · K^{-1}
Type IIa	2100 W · m^{-1} · K^{-1}
Resistivity (20°C or 68°F)	
Type Ia and IIa	$>10^{14}$ Ωm
Type IIb	0.1–10 Ωm

only 85 million years ago. Diamond-bearing kimberlites occur in South Africa, Botswana, Angola, Sierra Leone, Guinea, Tanzania, Brazil, Venezuela, the United States, Canada, Russia, Siberia, China, India and Australia. Only lamproites in Western Australia and one in Arkansas in the United States are diamond bearing.

Alluvial deposits. Kimberlite and lamproite are weathered and eroded by normal geological processes, and the diamonds are released. Because diamond is hard, chemically inert, and dense, it is carried into and along the river drainage systems. Diamonds, together with other minerals, eventually become trapped in alluvial or placer deposits. The diamonds in these secondary deposits may have traveled hundreds of miles, over millions of years, from the primary-source kimberlite. Typical alluvial diamond-bearing deposits occur along the Vaal and Orange rivers in South Africa, and in Angola, Zaire, the Central African Republic, Ghana, Sierra Leone, and Guinea. In South America diamonds are mined from alluvial deposits in Brazil, Guyana, and Venezuela. The historic gem diamonds recorded between the twelfth and seventeenth centuries were obtained from alluvial deposits in India. China, Kalimantan, Russia, and Western Australia are other sources for alluvial diamonds.

A unique type of alluvial deposit is the diamond-bearing marine gravel off the west coast of South Africa and Namibia. The diamonds in these deposits were transported by the Orange River from the interior of South Africa to the coast, where they have been redistributed by wave action and other marine processes. Generally, transport in a high-energy environment, such as that in a river or surf zone, results in the diamonds with serious flaws, defects, and cracks being destroyed, with only the better-quality diamonds surviving. The ultimate manifestations of this process are the marine deposits where 95% of the diamonds are of gem quality. In contrast, in a primary-source kimberlite or lamproite, the percentage of gem diamonds is much less; for example, in the Argyle mine, Western Australia, only about 10% of the diamonds are of gem quality.

Meteorites. Minute blackish crystals and small irregular polycrystalline masses (less than 0.04 in. or 1 mm) of diamond have been recovered from certain meteorites, for example, the Canyon Diablo iron meteorite which formed Meteor Crater, Arizona. There are varying opinions as to whether or not the diamonds were formed as a result of collision of asteroid-size bodies in space or as a result of impact with the Earth's surface. The latter idea is generally held. The ultrahigh-pressure form of carbon, lonsdaleite (hexagonal diamond), is also found associated with meteorites and impact craters. Lonsdaleite is often intergrown with polycrystalline masses of diamond (carbonado or framesite).

Mining and recovery. The major production of diamond is from the primary sources in South Africa, Botswana, Zaire, Australia, and Siberia, with minor amounts coming from Tanzania and China. The diamond mines in all these countries are based on kimberlite, except for the Argyle mine in Australia, where the source rock is lamproite. Although Botswana and South Africa produce the most gem diamonds, as well as Russia whose production is difficult to assess, the major worldwide production is from the Argyle mine in Australia, albeit mostly industrial diamonds. Diamond production from secondary (alluvial or placer) deposits, apart from the extensive mining of the marine gravels off the west coast of southern Africa, is relatively small compared to the output from mines based on kimberlite and lamproite. Alluvial deposits in Guinea, Ghana, Russia, and Australia are mined by large companies or government agencies. All other alluvial diamond deposits are worked by small local groups or individual miners. *See* PLACER MINING.

In small alluvial deposits diamonds are recovered by manual methods, which include sieving, jigging, and then searching the residue by eye. In the larger, more mechanized alluvial deposits, techniques similar to those of the major primary-source diamond mines are used. Diamond mines may be either open-pit or underground, although normally a new primary deposit is first mined as an open pit and later by underground development. Only the mines in South Africa (the Premier, Finsch, and Kimberley mines) are underground mines, although some mines in Siberia have been converted from open-pit to underground activity.

On average, diamond is 1 part per 20 million parts of waste rock. In diamond mining a value denoting the number of carats per 100 metric tons of rock is used to indicate the grade. At Kimberley the grade is about 80 carats/100 metric tons; at Mbuji Mayi, Zaire, the grade is roughly 200 carats/100 metric tons; at the Argyle mine in Western Australia, grades up to 400 carats/100 metric tons have been recorded. Nevertheless, the economics of diamond mining does not rest solely upon grade; factors such as the size and quality of the diamonds are equally important. *See* CARAT.

The mined rock is first crushed and screened to various sizes and then passed through heavy-media separation equipment. The intention is to separate the light fraction (less than 1.6 oz/in.^{3} or 2.7 g/cm^3), which is waste, from the heavy minerals, including diamond. This heavy fraction is referred to as the concentrate. Prior to the 1970s, much of this primary separation was done in rotary washing pans. In these pans a thick muddy liquid is kept at a density of about 0.41 oz/in.^{3} (1.25 g/cm^3). The lighter material floats to the surface, whereas diamond and other heavy minerals concentrate at the bottom of the pan and are periodically removed. Modern heavy-media separation equipment works on a similar principle, but it has more efficient control of the separation media.

The concentrate then flows with water over a vibrating table, or conveyor belt, covered with grease. Diamond has an affinity for grease and sticks to it, whereas other minerals pass over the grease with the water and are rejected. Although grease tables are used in some mines or as secondary recovery

methods, most modern plants use a method based on x-rays to separate diamonds from the waste material in the concentrate. When diamonds are irradiated by x-rays they luminesce, whereas most other minerals in the concentrate do not. The concentrate, in a narrow stream, passes through a beam of x-rays. A light-sensitive module senses the luminescence of the diamond and sends an impulse, usually to an air jet, to blow the diamond out of the stream and into a sealed container. This method is over 97% reliable and is referred to as the sortex method. Both heavy-media separation and sortex recovery may be aided by rotary pans, heavy-media cyclones, grease tables, electrostatic separators, and skin flotation. The last method is usually used to recover diamonds smaller than 0.04 in. (1 mm) in diameter. After recovery the rough diamonds are cleaned in acid and sorted for size, shape, quality, and color.

Cutting and polishing. Rough diamonds occur in a variety of shapes, including octahedra, dodecahedra (**Fig. 1**), twinned octahedra (macle), and broken or cleavage fragments. The largest diamond found, the Cullinan, was a cleavage fragment. After the rough is sorted into cuttable (gem and near gem) and industrial stones, the decision is made as to how a specific diamond will be shaped and made into a polished gem. The cutting and polishing process can result in the loss of as much as 60% of the original diamond.

The initial step is to alter the shape of the rough diamond so that it approximates more closely that of the finished gem. This may be done by cleaving or sawing. Prior to either of these actions being taken, the diamond cutter will determine the grain of the stone. The grain reflects specific crystallographic directions in the diamond along which cleaving and polishing can be more readily accomplished. The direction of the grain is often observed from natural lines or patterns on various crystal faces. In cleaving a stone, the cleaver will follow the octahedral grain. A small nick, known as the kerf, is made along the direction of cleavage in the diamond by scratching another diamond along the direction to be cleaved. A thin steel blade is then placed in the kerf and tapped; if done correctly the diamond cleaves perfectly.

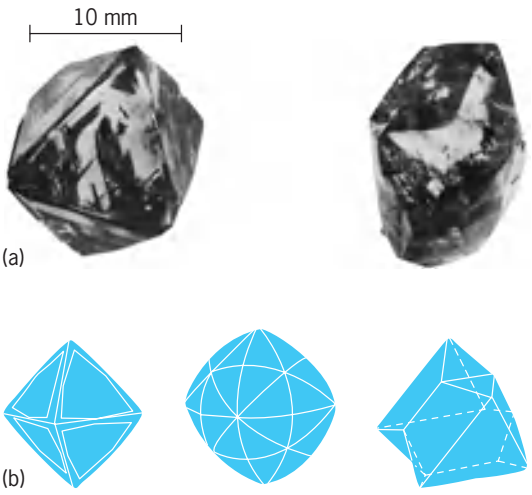

Fig. 1. Diamond. (*a*) Crystals from Kimberley, South Africa (*American Museum of Natural History specimens*). (*b*) Crystal habits (*after C. Klein and C. S. Hurlbut, Jr., Manual of Mineralogy, 21st ed., John Wiley and Sons, 1993*).

Instead of cleaving, it may be necessary to saw a diamond. The saw consists of a thin phosphor bronze disk whose edge is charged with paste of fine diamond powder and olive oil. The diamond is mounted in a holder and placed on the upper part of the revolving cutting disk. Depending upon the size and nature of the diamond, it may take hours or even days to complete the sawing. Instead of sawing, in modern practice many diamonds are sliced with a laser beam. This technique has the advantage of speed, but more importantly it allows the cutting of stones with irregular areas of crystal growth (naats or knots) that are difficult or impossible to saw.

The next step in the process is bruting. The bruter (also cutler or cutter) uses a second diamond as a tool in order to round off the corners and edges of a rough stone or of a cleaved or sawn diamond. The aim is to prepare the stone to a rounded shape upon which the facets of the final gem can be ground.

The last stages are grinding and polishing facets on the diamond. These stages are done on a scaife, a horizontal cast-iron wheel dressed with a fine diamond powder and olive oil. The diamond is placed in a dop, a holder that can be adjusted to enable different faces to be ground and polished. Grinding is the initial coarse preparation of the facets, which are then polished with a finer grain size of diamond paste. When completed, a brilliant-cut diamond, the most common shape, consists of 58 symmetrical facets of different sizes and relative angles. Automation is increasingly used in the polishing industry, and there are a number of machines available that can be programmed to automatically grind and polish the facets simultaneously on several diamonds. For the most part, the automation is used on smaller stones; the larger and more important are polished by hand. Prior to about 1940, most cutting and polishing was done in Amsterdam and Antwerp. Since then diamond polishing centers have opened in London, New York, and Tel Aviv and in Russia, India, southern Africa, and China and other localities in the Far East.

Polished diamonds are graded on the basis of the 4 C's—carat, cut, clarity, and color. The carat is the unit of weight in the diamond industry and is standardized as 0.2 gram (0.0071 oz or 200 milligrams) and is divided into 100 points. Thus a 10-point diamond weighs 0.1 ct (0.00071 oz or 20 mg). The largest diamond, the Cullinan, weighed about 3000 ct (4.27 oz or 600 g) and was the size of an average human fist. The grading cut is based on how well the facets and the shape of a polished diamond compare to a standard model.

In some instances, in order to save weight or to hide blemishes, the diamond may be cut and polished to be thinner or thicker than the standard proportions. The clarity of a polished diamond depends upon the presence or absence of defects such as inclusions, cleavages, and other natural flaws.

A diamond with no such defects and having top clarity is referred to as being flawless.

Diamonds, although commonly considered to be mostly colorless, actually exist in all the colors of the rainbow. Colored stones are known as fancies, and if of excellent uniform color they are most desirable. Perhaps the most famous colored diamond is the deep-blue Hope diamond (44.5 ct), displayed in the museum of the Smithsonian Institution. The color of diamonds can be artificially changed by various types of irradiation; for example, neutrons, protons, electrons, alpha particles, gamma particles, and radioactive isotopes. Irradiation, followed by heat treatment, can turn a pale yellow, relatively worthless diamond into a pale to dark green stone, which if it were a natural color would be a prized stone. Unfortunately, it is not always possible to distinguish between an enhanced and naturally colored diamond. *See* GEM.

Synthesis. Diamonds were first synthesized in Sweden in 1953. However, the information was kept secret until after the General Electric Company in the United States was issued a patent. These early experiments used the principle that carbon dissolves in the transition elements of groups 8–10 (such as iron or nickel). At high pressures (50–60 kilobars or 5–6 gigapascals) and temperatures (1500°C or 2700°F), the dissolved carbon nucleates and crystallizes as diamond. Direct conversion of graphite to diamond was achieved in 1961 in shock-wave experiments in which transient high pressures in excess of 300 kb (30 GPa) and temperatures of about 1100°C (2000°F) existed. A similar natural phenomenon is believed to be the origin of diamonds in meteorites.

Unlike imitation diamonds, which consist of nondiamond material such as glass, cubic zirconia, or strontium titanate, synthetic diamond is identical with the natural mineral. The properties (for example, crystal structure or density) are similar, although minor differences do exist because of the different processes of synthetic and natural growth. Thus synthetic diamonds often have a pale yellow color due to the incorporation of atmospheric nitrogen in the synthesis chamber; consequently most synthetic diamonds are type Ib. Synthetic diamonds may also contain metallic inclusions that represent the flux from which they were grown; common are tiny inclusions of nickel. Furthermore, although diamonds can be synthesized with a variety of shapes, most are cubo-octahedral, a form virtually absent in natural stones. Synthetic diamonds generally are not large; most are produced in sizes below 0.004 in. (0.1 mm). These are used extensively as grit for industrial grinding purposes. Colorless gem-quality diamond can also be synthesized, but the cost of synthesis has proved to be greater than the cost of the natural product. Synthetic gem crystals greater than 11 carats have been manufactured. It is possible to differentiate between such synthetic and natural diamonds.

In contrast to growing diamond at high pressures and temperatures, diamond can also be synthesized at pressures below 1 atm (10^5 pascals) and at temperatures of around 1000°C (1900°F). In this process, carbon is deposited from a carbon-rich vapor as thin polycrystalline films of diamond on a substrate of diamond, silicon, glass, or some other suitable material. The initial experiments for this type of synthesis date from about 1962, but detailed work in the Soviet Union in the late 1960s led to intensive research in Japan and then worldwide. Diamond films can be grown in several ways. For example, diamond crystals up to 0.02 in. (0.5 mm) in size can be formed from a mixture of methane and hydrogen at about 50 torr (6.7 kPa) pressure and 1000°C (1900°F) on a silicon substrate. This method is known as thermally induced chemical vapor deposition, but other techniques may be used, including plasma chemical vapor ion deposition and electron-beam deposition. The diamond films display properties similar to those of natural diamond and have similar hardness and thermal conductivity, both significant properties for the uses of diamond films.

Applications. Diamond, apart from its use as a gem, has numerous applications in industry, and it is designated a strategic mineral. Many of the uses of natural and synthetic diamond are equivalent. Originally, natural diamond, including boart, carbonado, and framesite, was crushed to various sizes of powder and used as grinding and polishing agents for glasses, ceramics, and nonferrous metals. Diamonds, as single crystals or powders, are also bonded in metal drills and bits. Small drills are used in applications such as dental work; large drills are used in drilling for oil and other minerals. Diamond-impregnated wheels are used for cutting many hard materials, including concrete and dimension stone for architectural purposes. Synthetic diamond is sometimes preferred for various uses, as it is grown to the specific grain size rather than crushed as in the case of natural diamond. Crushing results in inherent cracks or fractures in the individual grains that decrease the life use of the material.

Diamonds are used in eye surgery, and also as heat sinks and semiconductors in the electronics industry. Diamond films (**Fig. 2**) have potential uses as scratchproof coatings on optical lenses, compact

Fig. 2. Scanning electron micrograph of a diamond film deposited by microwave-induced plasma, showing the crystalline morphology.

discs, and even on nondiamond jewelry; bearings in machines; heat sinks and semiconductors in electronics; and general inert coatings or surfaces in areas of high chemical corrosion. Natural diamond has also been used as optical windows in spacecraft. Henry O. A. Meyer

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Diamond thin films

Very thin coatings that are applied to other materials to provide them with special physical or chemical properties unique to diamond. While crystalline diamond is extremely expensive, many of the useful properties of diamond can be obtained in a cost-effective manner by coating a substrate of interest with a polycrystalline film consisting of tiny diamond crystals that form a continuous film, often 1 micrometer or less in thickness. Diamond thin films can be grown on a variety of thicker materials that provide mechanical support and possibly electrical or thermal contact.

Fabrication. Diamond and graphite are different forms (allotropes) of pure carbon. Because graphite is the more thermodynamically stable form, growth of diamond requires special procedures. The formation of diamond thin films consists of two distinct steps: nucleation and growth. Nucleation is the initial formation of very small diamond crystallites. Growth refers to the set of chemical and physical processes that make these crystallites grow larger and ultimately form a continuous film consisting of many tiny, interlocked crystals (**Fig. 1**). The initial nucleation of diamond must usually be enhanced. Commonly used methods include "seeding" the growth mechanically by abrading the substrate with very fine diamond powder or by ion bombardment. *See* DIAMOND; GRAPHITE.

Growth of diamond films is typically performed via a chemical vapor deposition process in which methyl radicals ($CH_3\cdot$) are first created by decomposing methane (CH_4) gas, removing one hydrogen atom and its associated electron (H·) from a molecule of methane via $CH_4 \rightarrow CH_3\cdot + H\cdot$. The methyl radicals then deposit on the solid surface. While both diamond and graphite can be formed, the selective growth of diamond is achieved by also having a large excess of hydrogen present; hydrogen atoms preferentially "etch away" any graphite that forms, leaving behind a pure diamond film. The energy needed to break apart the CH_4 and H_2 molecules can come from several different sources. One common way is to use microwaves to ignite a plasma in a mixture of methane and hydrogen (typically ~5% methane in 95% hydrogen). The deposition of methyl radicals onto the surface leads to growth of tiny diamond crystals, which grow in size and ultimately form a continuous film consisting of a tightly interlocked network of diamond crystals. The boundaries between these crystals have some influence on the properties, and the size and shape of the crystals can be controlled during growth to produce films in which specific properties are optimized (**Fig. 2**). Diamond growth rates are slow, typically 1–10 μm/h for high-quality films; faster growth rates lead to higher incorporation of graphite into the film and a degradation of most properties. *See* CHEMICAL VAPOR DEPOSITION; PLASMA (PHYSICS).

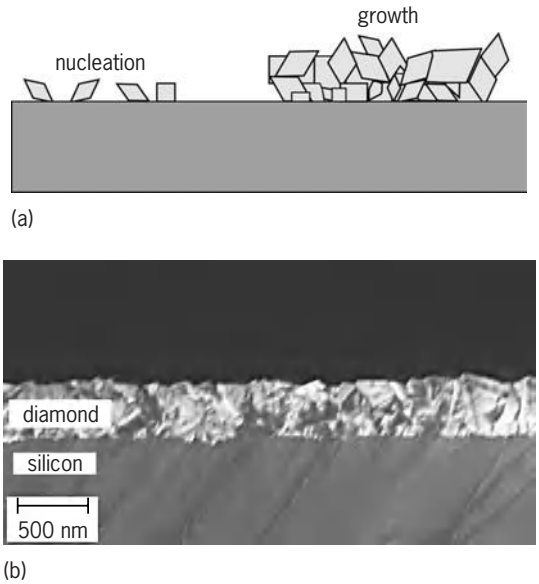

Fig. 1. Diamond thin films. (*a*) Nucleation and growth of a diamond thin film. (*b*) Scanning electron microscope side view image of a 400-nm-thick diamond thin film grown on a silicon substrate. (*Wensha Yang, University of Wisconsin-Madison*)

Properties. Much of the interest in diamond stems from its many extraordinary physical and chemical properties.

Extraordinary hardness. Diamond is the hardest known naturally occurring material.

Low friction. Diamond surfaces exhibit quite low friction.

High thermal conductivity. At room temperature, diamond conducts heat five times better than copper.

Optical transparency. Diamond transmits a wide range of wavelengths, including microwave, infrared, and ultraviolet, making it useful for specialized optics for satellites, spacecraft, and specialized measurements.

High index of refraction. The high index of refraction of 2.41 enhances reflections and contributes to diamond's sparkle; for example, in jewelry.

Electrical properties. Diamond is an insulator but can be "doped" (for example, with boron atoms) during the growth process to give it semiconducting properties.

Chemical inertness. Diamond is stable in aggressive chemical and physical environments.

Biocompatibility. Diamond and diamondlike films are used as coatings on biomedical implants.

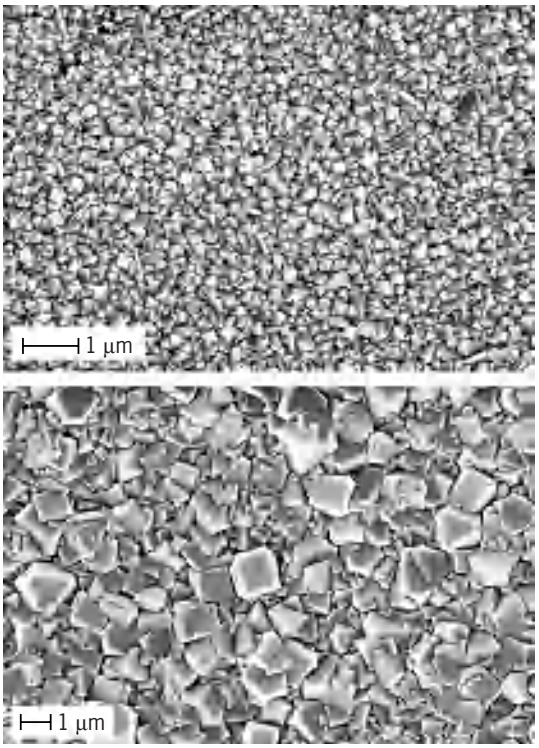

Fig. 2. Scanning electron microscope images of two diamond films, showing the individual diamond crystallites that agglomerate into a continuous film. (*a*) Very small diamond grains. (*b*) Film with larger diamond crystals. (*Bin Sun, University of Wisconsin-Madison*)

Applications. The extreme mechanical, optical, and chemical properties of diamond make it an attractive material for use in a variety of technological applications. The ability to grow diamond as a thin film provides a way to take advantage of these properties in a practical, cost-effective manner. Diamond is very useful as a thin coating on machine tools such as drills, saws, and other cutting tools. Like silicon, diamond's electrical conductivity can be changed at will, making it potentially useful for specialized applications in electronics. Diamond's high thermal conductivity makes it useful for carrying heat away from high-power electrical devices such as diode lasers. In addition, diamond has been receiving much attention because of its chemical stability, making it an attractive substrate for applications such as environmental monitoring and biological sensing.

Robert J. Hamers

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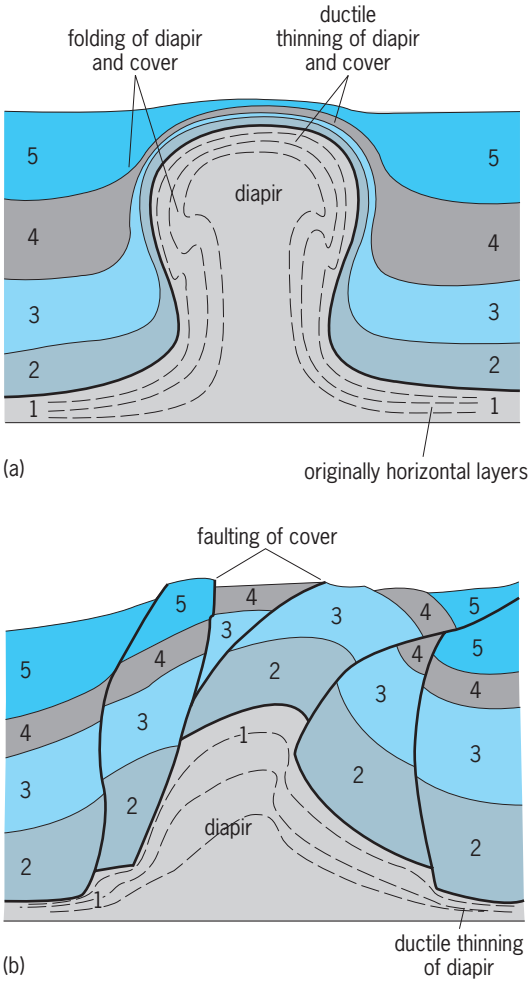
Diapensiales

An order of flowering plants, division Magnoliophyta (Angiospermae), in the subclass Dilleniidae of the class Magnoliopsida (dicotyledons). The order has only a single family, the Diapensiaceae, with about 18 species of herbs and dwarf shrubs in temperate and arctic regions of the Northern Hemisphere. Within its subclass the order is marked by sympetalous flowers (flowers with the petals joined to each other by their margins, at least toward the base, forming a basal tube, cup, or saucer), separate pollen grains, and ovules that have a single integument (unitegmic) and are tenuinucellate (with the nucellus consisting of a single layer of cells covering the micropylar end of the embryo sac). *Shortia* and *Galax* genera are sometimes cultivated. *See* DILLENIIDAE; FLOWER; MAGNOLIOPHYTA; PLANT KINGDOM.

Arthur Cronquist; T. M. Barkley

Diapir

A buoyant mass of ductile rock or sediment that has pierced, or appears to have pierced, overlying rock, known as overburden. The overburden can yield by ductile processes or by brittle faulting (see **illus.**). Diapirs form by lateral and vertical intrusion of buoyant or nonbuoyant rock. Emplacement of a diapir, known as diapirism, involves piercement of overlying rocks.



Vertical sections showing diapirism by (*a*) ductile folding of cover and (*b*) faulting of cover. Number 1 indicates the bedding of the diapir, and 2–5 identify cover layers.

Diapirs are composed of salt, gypsum, shale, mud, sand, peat, coal, limestone, ice, serpentinite, granite, gneiss, and migmatite. Salt diapirs are typically several miles wide and high, but sheetlike varieties that have spread or coalesced laterally can be as much as 200 mi (300 km) wide. This type of diapir is economically important. Gneiss and migmatite diapirs are typically 6–13 mi (10–20 km) wide but locally approach 60 mi (100 km) in width. Diapirs doming the sea floor form large islands of salt in the Persian Gulf and islands of shale in the Caspian and Banda seas. *See* GNEISS; GRANITE; MIGMATITE.

Over geologic time, solid rock can flow slowly by creep. Like glacial ice, salt can flow at the surface, but for gneiss to creep requires burial to a depth of at least 9 mi (15 km), where temperatures of 1100°F (600°C) have softened the rock. Diapirism is driven by gravitational potential energy released by the sinking of denser overburden around the diapir. Either the overburden becomes denser by compaction or metamorphism, or the diapiric rock becomes less dense by heating or infusion of water. Hot diapirs rise as part of convection systems. Even diapirs denser than their overburden can rise in grabens or in the cores of folds formed by forces stretching or compressing the crust because of plate motions or the effects of gravity on sloping continental margins. Linear loads can cause diapirs to be shaped like irregular walls, instead of plugs, which are more common. Where updomed roof strata are eroded, salt diapirs can extrude through the eroded orifice. Such diapirs grow higher by remaining near the sedimentation surface while their bases sink with the subsiding basin floor, a process known as downbuilding. *See* GRABEN; SALT DOME.

M. P. A. Jackson

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Diapsida

A group of reptiles characterized by two pairs of lateral skull openings (fenestrae) and a suborbital fenestra in the palate below the eye. This group includes living and extinct lizards, snakes, tuatara, and crocodiles, as well as a range of extinct Late Paleozoic and Mesozoic lineages including rhynchosaurs, protorosaurs, younginiforms, choristoderes, dinosaurs, and pterosaurs. In addition, most workers would now include the so-called parapsid or euryapsid reptiles (defined as possessing a single upper-skull fenestra; **Fig. 1***c*), with the view that the ancestral lower temporal fenestra was secondarily reduced or closed in the aquatic nothosaurs, plesiosaurs, pliosaurs, ichthyosaurs, and placodonts. Perhaps more surprisingly, some authors also consider the anapsid (no skull fenestrae) turtles to be diapsid derivatives, and this idea, while still contentious, has received support from many molecular phylogenetic studies, which use deoxyribonucleic acid (DNA) sequence data to determine evolutionary relationships. Birds evolved from bipedal theropod dinosaurs and are also part of the diapsid radiation, although older, traditional classifications place them in a separate class, Aves. *See* ANAPSIDA; ANIMAL EVOLUTION; AVES; DINOSAUR.

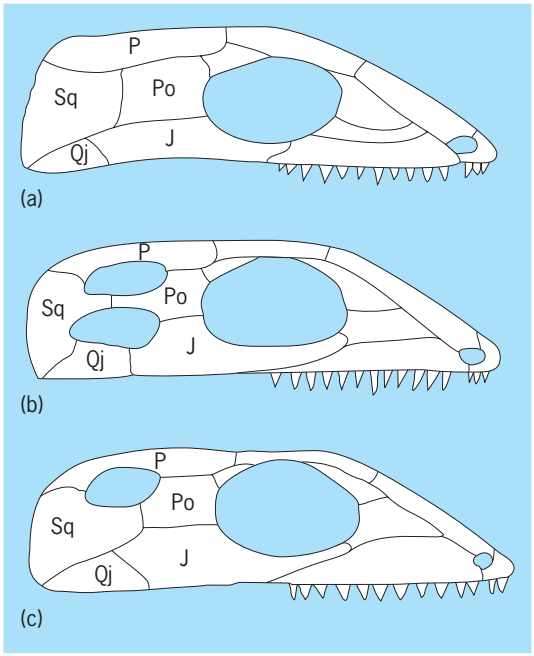

Fig. 1. Series of three skulls in lateral view: (*a*) anapsid skull condition, (*b*) diapsid skull condition, and (*c*) parapsid/euryapsid skull condition. J, jugal; P, parietal; Po, postorbital; Qj, quadratojugal; Sq, squamosal.

Skull morphology. The upper and lower fenestrae that characterize diapsids lie in the temporal region of the skull between the orbit and the ear region. This region houses the main jaw adductor muscles. The two openings arise in the position of three- and four-point junctions between skull bones in the ancestral anapsid skull (Fig. 1*a*), rather like the fontanelles in the neonatal human skull. Thus the upper fenestra usually lies between the parietal, postorbital, and squamosal bones (Fig. 1*b*), and the lower fenestra lies between the postorbital, jugal, quadratojugal (where present), and squamosal bones (Fig. 1*b*). This leaves a postorbital bar between the fenestrae and the orbit, an upper temporal bar between the upper and lower fenestrae, and a lower temporal bar below the lower fenestra. Classically the lower temporal bar is complete, but it is missing or incomplete in many diapsid lineages, either through multiple losses (for example, squamates, rhynchocephalians, protorosaurs) or as a primitive state, with early loss being followed by the independent redevelopment of a complete bar on several occasions (Tuatara, archosaurs). In snakes and some lizards, the upper temporal bar and postorbital bars are also lost as an adaptation to skull

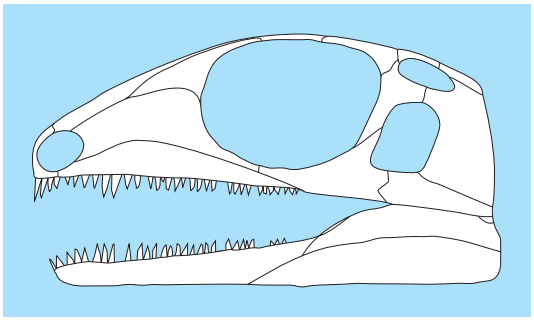

Fig. 2. Lateral view of the skull of the basal diapsid *Petrolacosaurus*. (*After R. R. Reisz, A diapsid reptile from the Pennsylvanian of Kansas, Univ. Kansas Publ. Mus. Nat. Hist., 7:1–74, 1981*)

kinesis (the ability of parts of the skull to move or flex relative to each other).

Evolution. The earliest adequately documented diapsid is *Petrolacosaurus* from the Upper Carboniferous (Pennsylvanian) of Kansas [305 million years ago (mya); **Fig. 2**]. It was lightly built and about 8 in. (20 cm) in length, with a relatively long neck and long distal limb elements. Apart from these differences in proportions and the presence of the diagnostic temporal openings, the skeleton of *Petrolacosaurus* strongly resembles that of more primitive amniotes such as *Hylonomus* (Upper Carboniferous, Canada; 310 mya). The Lower Permian *Araeoscelis*, also from North America (Texas and Oklahoma, 280 mya), is now considered to be a close relative of *Petrolacosaurus*, although it was originally classified as a parapsid due to the absence of a lower temporal fenestra (probably secondarily closed).

Diapsids remain poorly documented throughout the Late Paleozoic, an era dominated by the synapsid ancestors of mammals. Nonetheless, Upper Permian and Lower Triassic rocks chart the early stages of a major radiation. With the decline of the nonmammalian synapsids beginning at the end of the Permian and accelerating through the Triassic, diapsids diversified and came to dominate Mesozoic ecosystems—land, sea, and air. They remain highly successful today although they have been somewhat eclipsed by the second major synapsid radiation (of modern mammals). Some Permian and early Mesozoic lineages, including coelurosauravids (Upper Permian gliders from Europe and Madagascar, 255 mya), younginiforms (small terrestrial and aquatic reptiles from the Late Permian and Triassic of Africa and Madagascar, 255–245 mya), and thalattosaurs (long-snouted marine reptiles from the Triassic of North America, Europe, and China, 235–225 mya) may have branched off the diapsid stem in the early stages of the radiation. Others are more difficult to place. Specialization toward an aquatic lifestyle affects characters of the limbs, girdles, and vertebral column that would otherwise be important for classification. As a result, the phylogenetic positions of groups such as the placodonts, plesiosaurs, choristoderes, and ichthyosaurs remain problematic. *See* SYNAPSIDA.

The remaining advanced diapsids are grouped into one of two major clades, Lepidosauromorpha and Archosauromorpha, represented in the modern fauna by the lepidosaurs (lizards, snakes, amphisbaenians, tuatara) on the one hand and by the archosaurs (crocodiles and birds) on the other (**Fig. 3**). The two major diapsid groups are distinguished by skull morphology and by divergent specializations of the vertebral column, limbs, and girdles affecting posture and locomotion. *See* ICHTHYOPTERYGIA; PLACODONTIA; PLESIOSAURIA.

Lepidosauromorpha. Most lepidosauromorphs have a lightly built skull with an incomplete lower temporal bar, large eyes, and an ear region structured to support a large eardrum. They retain an essentially sprawling posture and rely on lateral undulation of the trunk in locomotion, but they developed specialized vertebral joints for added stability, and more sophisticated knee, ankle, and foot joints to improve movement over a variety of substrates. For small lepidosauromorphs, the retention of a sprawling posture was not a serious handicap, and most can run and climb on a variety of surfaces. Large-bodied lepidosauromorphs are rare and most have either been aquatic (for example, the Late Cretaceous marine mosasaurs, 93–65 mya) or secretive and slow moving (for example, heavy constricting snakes such as the python and anaconda). The exceptions are the monitor lizards of the family Varanidae, a terrestrial lineage that has repeatedly achieved large size (for example, the living Indonesian Komodo dragon and the extinct Australian *Megalania*, about 1.6 mya–40,000 years ago), but in regions devoid of large mammalian predators.

Basal lepidosauromorphs probably separated from the ancestors of archosaurs in the Early or Middle Permian, but they remain poorly represented in Early Mesozoic deposits. The record of basal lepidosauromorphs (that is, those classified outside Lepidosauria) includes the fragmentary Early Triassic *Paliguana* from South Africa (250 mya), as yet unnamed fossils from the Early Triassic of Poland, and the highly specialized long-ribbed gliding kuehneosaurs (*Kuehneosaurus*, *Icarosaurus*, 210–200 mya) from the Upper Triassic of Britain and North America. In the past, both *Paliguana* and kuehneosaurs were considered to be early lizards, but this was based on superficial features such as small size and a lightly built skull. True lepidosaurs are linked by a large suite of characters. The ancestors of lizards and snakes (Squamata) apparently diverged from those of tuatara (Rhynchocephalia) early in the Triassic but they do not appear in the fossil record until later. The first rhynchocephalians are recorded from the beginning of the Late Triassic (Germany, 220 mya), but the derived skull and dental characters of these early representatives show that they had a long unrecorded history before this. Similarly, although squamates are first known from the Early/Middle Jurassic (India, Britain, Central Asia, 160 mya), their level of diversity at that time is indicative of a pre-Jurassic history. *See* LEPIDOSAURIA; SQUAMATA.

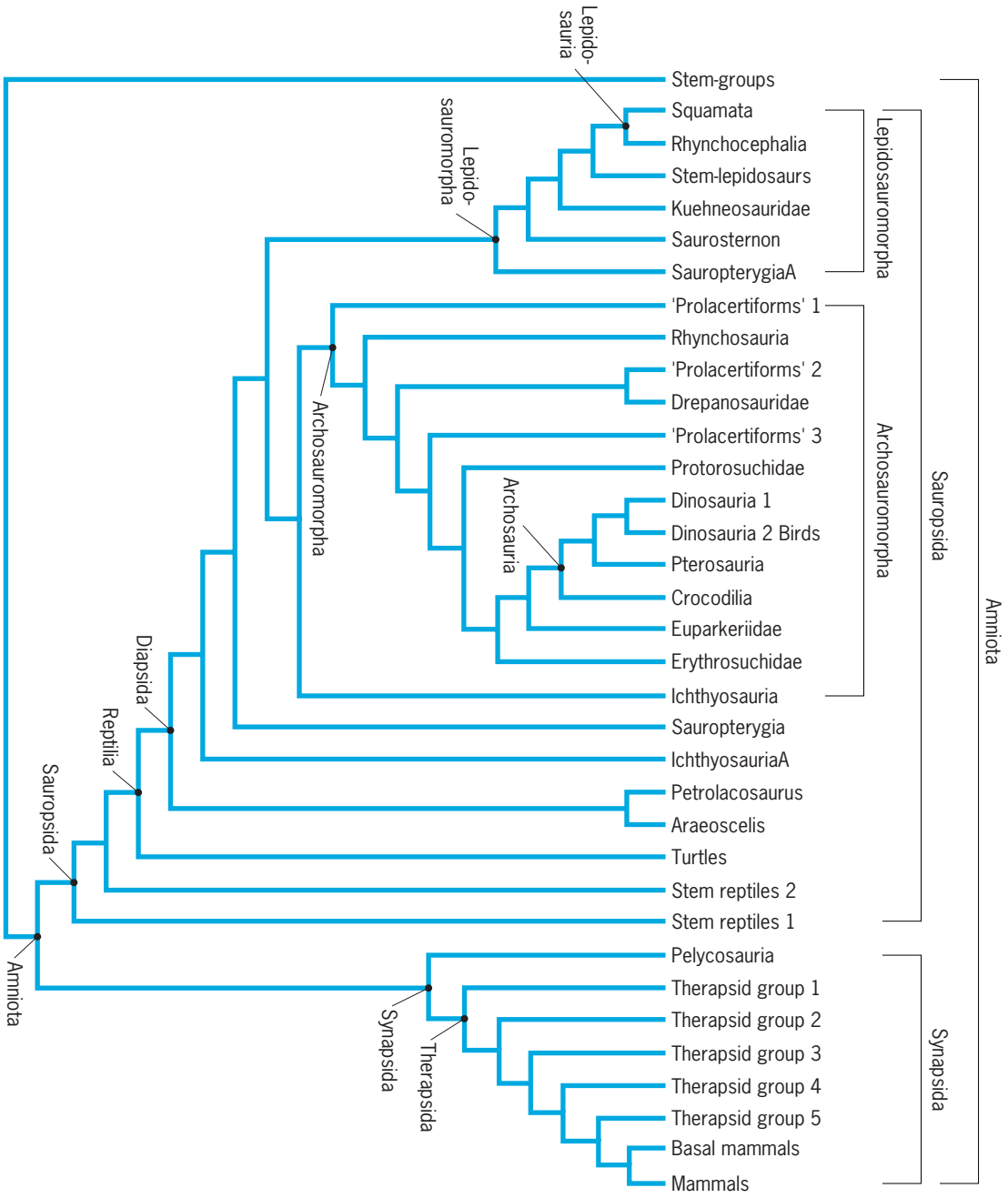

Fig. 3. Phylogenetic tree showing relationships of the main groups of amniotes, including the primary division into Sauropsida and Synapsida, and the division of Diapsida into Lepidosauromorpha and Archosauromorpha. Sauropterygia and Ichthyosauria each appear in two places in the tree because there is some debate as to their relationships. Therapsid numbers 1, 2, etc., denote different therapsid groups; however, these are purely arbitrary numbers, as the tree is really much more complex with many more branches at each level. The same applies to stem reptiles 1 and 2.

Archosauromorpha. Primitive archosauromorphs were widely distributed during the Upper Permian and Triassic and include the long-necked protorosaurs (also known as prolacertiforms), the apparently herbivorous rhynchosaurs and trilophosaurs, the arboreal or swimming drepanosaurs (Late Triassic, Italy and North America, 225 mya), and perhaps the gliding *Sharovipteryx* (Upper Triassic, Kyrgyzstan, 225 mya). Primitive archosauromorphs such as the protorosaurs and rhynchosaurs were essentially improved sprawlers, but derived genera brought the limbs under the body and coupled a more rigid trunk region with the development of a parasagittal gait. This reached its fullest expression in dinosaurs and birds that resemble mammals in having an energetically efficient upright stance and a vertebral column that flexes primarily in the dorsoventral plane. Inevitably, this allowed an increase in body size, as seen in some of the massive herbivorous dinosaurs of the Jurassic and Cretaceous. *See* ARCHOSAURIA; PROTOROSAURIA; RHYNCHOSAURIA.

More derived, archosaur relatives (grouped with

Archosauria in the more inclusive group Archosauriformes) include the primitive protorosuchians (Upper Permian to Lower Triassic, 255–245 mya, South Africa, Australia, Russia), the large erythrosuchids (Early to Middle Triassic, 250–230 mya, South Africa, China, Russia), and the more gracile euparkeriids (Early Triassic, 250 mya, South Africa, Russia, Poland).

Within Archosauria, the lineage leading to crocodiles was distinct by the Middle Triassic, as were the immediate ancestors of the dinosaurs and the pterosaurs (Fig. 3). Dinosaurs began to radiate during the Upper Triassic, and by the Upper Jurassic most of the major groups (including birds) had appeared. At the end of the Mesozoic, all the remaining nonavian dinosaurs became extinct, along with the pterosaurs, mosasaurs, and plesiosaurs. Crocodiles, most lizards, snakes, tuatara and birds all survived (as did turtles, amphibians, and many mammals). *See* REPTILIA.

Susan E. Evans

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Diarrhea

The passage of loose or watery stools (greater than 8 oz or 200 g per 24 h), usually at more frequent than normal intervals. The two principal mechanisms responsible for diarrhea are the rise in osmotic pressure of fecal contents, diminishing the absorption of water by the intestine (osmotic diarrhea); and the increase in electrolyte secretion into the intestinal lumen by its epithelial cells, leading to an increased flow of water from the intestinal mucosa into the lumen (secretory diarrhea).

Pathophysiology. Ingestion of salts containing magnesium, sulfate, or phosphate ions can bring about osmotic diarrhea since these ions are poorly absorbed by epithelial cells of the intestinal lumen; consequently, water absorption is diminished. Similarly, diarrhea may also result from the impaired absorption of disaccharides or monosaccharides. For example, individuals with a deficiency of the intestinal disaccharidase enzyme lactase have an impaired ability to split dietary lactose into glucose and galactose. As a result, such individuals can have watery diarrhea after ingesting milk, since lactose molecules remain in the intestinal lumen and inhibit water absorption.

Just as the increased transport of electrolytes and water into the intestinal lumen can bring about secretory diarrhea, the diminished absorption of electrolytes and water can also cause it and contribute to clinical manifestations in patients with secretory diarrhea. In some diseases that are characterized by secretory diarrhea (for example, cholera), fecal water can exceed several quarts per 24 h. Cholera is caused by the bacterium *Vibrio cholerae*, which colonizes the small intestine and produces a protein toxin responsible for secretory diarrhea. These toxic molecules stimulate the production of the enzyme adenylate cyclase in the epithelial cells of the small intestine, which leads to the conversion of adenosine triphosphate (ATP) to cyclic adenosine monophosphate (AMP). Cyclic AMP induces secretion of chloride ions and inhibits absorption of sodium ions by epithelial cells of the small intestine. These events lead to water secretion, impaired water absorption, and profuse watery diarrhea. Strains of *Escherichia coli* that cause the diarrhea common to travelers produce a toxin having a similar mode of action to that of cholera toxin. *See* CHOLERA.

Bacillary dysentery, which also is characterized by secretory diarrhea, is caused by bacteria of the genus *Shigella*, which invade epithelial cells of the colon and release a toxin that kills the cells by blocking their ability to synthesize proteins. Individuals with bacillary dysentery have ulcerations of the colon and excrete blood and mucus by way of the rectum. *See* BACILLARY DYSENTERY; FOOD POISONING.

Agents responsible for viral gastroenteritis include rotavirus and Norwalk virus. These viruses cause watery diarrhea, probably by colonizing and damaging epithelial cells of the small intestine, impairing their ability to absorb electrolytes and water. Protozoan parasites, including *Cryptosporidium* species and *Giardia lamblia*, cause diarrhea by mechanisms that are not well understood. *Cryptosporidium* infestation leads to profuse watery diarrhea in patients with acquired immune deficiency syndrome (AIDS). *See* MEDICAL PARASITOLOGY.

Impaired absorption of bile salts or long-chain fatty acids also causes secretory diarrhea. Nonabsorbed bile salts are deconjugated to free bile acids by bacteria in the colon, and colonic bacteria also hydroxylate nonabsorbed long-chain fatty acids. The bile acids and hydroxy fatty acids stimulate colonic secretion of electrolytes and water, which accounts for the secretory diarrhea.

Crohn's disease and ulcerative colitis are characterized by intestinal inflammation and diarrhea, with or without rectal bleeding. In these conditions diarrhea is due, at least partly, to impaired absorption of electrolytes and water, as a result of intestinal epithelial cell damage; however, the underlying cause is unknown. Diarrhea is also a feature of irritable bowel syndrome, which is a condition in which intestinal motility is increased and often appears to be a physical manifestation of anxiety or depression.

Identification, management, and treatment. If a bacterial infection or parasitic infestation is suspected, fecal microscopy and culture should be performed. Microscopy may reveal parasites (for example, *G. lamblia* cysts or *Cryptosporidium* oocysts), and culture may identify a bacterial pathogen. Where there is chronic diarrhea of unknown cause, some of the following tests may be informative: fecal fat

measurement (to test for malabsorption of fat in the intestine); endoscopic examination of the colon (to explore for a tumor or ulcerative colitis); barium radiography of the colon or small intestine, and mucosal biopsy of the small intestine (to identify diseases causing malabsorption); and examination of a fecal specimen for evidence of laxatives that a patient may surreptitiously be ingesting (for example, magnesium ions or phenolphthalein, present in many laxatives).

Management of diarrhea includes rehydration, and treatment of the causative disease. If a bacterial or protozoan pathogen can be identified as the cause, the person may benefit from treatment with an antimicrobial agent. However, not all bacterial infections of the intestine are eliminated by antimicrobial treatment, even if the bacterium is sensitive to the antimicrobial agent in the laboratory. Bismuth subsalicylate and drugs that reduce intestinal motility (for example, opiates and loperamide) are useful agents for the symptomatic treatment of diarrhea. *See* GASTROINTESTINAL TRACT DISORDERS.

Martin F. Heyworth

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Diastem

A break in the stratigraphic record produced by local erosion or nondeposition and representing a short interval of geologic time. The breaks in deposition are those that would occur within a particular sedimentary environment, rather than those associated with a major change in environment. *See* STRATIGRAPHY; UNCONFORMITY.

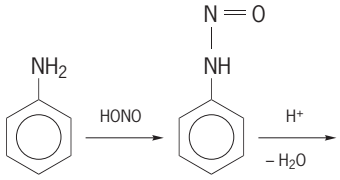
Diastems are not distinctive or even identifiable features. Their presence was deduced by J. Barrell in 1917. He recognized that deposition in modern environments is not continuous, so that many short intervals of geologic time must go unrecorded. Although the rate of deposition of an individual bed may be rapid, the presence of diastems results in a slow rate of accumulation for the stratigraphic sequence as a whole. Thus the aggregate time represented by strata in a given sequence may be much less than the time encompassed by the sequence.

Nondeposition can result either from excessive turbulence in the environment or from a lack in sediment supply. Discontinuity in sedimentation occurs on many scales. Short breaks, from seconds to days, are associated with migration of bedforms, variation in wave or current energy, and tidal cyclicity. Seasonal deposition occurs due to floods, storms, and plankton blooms. Sedimentation by major floods or hurricanes may occur on a scale of decades to centuries. In the deep ocean, the interval between successive turbidity current flows can be thousands of years. *See* MARINE SEDIMENTS; SEDIMENTOLOGY; TURBIDITY CURRENT.

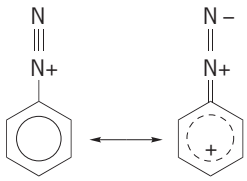
Charles W. Byers

Diazotization

The process by which an aromatic primary amine is converted to a diazonium compound. The preparation and reactions of diazonium salts were discovered in 1858 and were the basis of the synthetic dye industry and the development of other industrial chemistry in Europe. In diazotization, sodium nitrite is added to a solution of the amine in aqueous acid solution at 0–5°C (32–41°F). Reaction of the amine with nitrous acid gives a nitrosamine. Tautomerization and loss of water lead to the diazonium ion, which is stabilized by delocalization of the positive charge at the ortho and para carbon atoms of the ring [reaction (1)].



Nitrosamine



(1)

Diazonium ion

See AMINE; AROMATIC HYDROCARBON; DELOCALIZATION; TAUTOMERISM.

The overall reaction is simple and very general. Substituents of all types—alkyl, halogen, nitro, hydroxyl, sulfonic acid—can be present at any position. Heterocyclic amines such as aminothiazole or aminopyridines can also be diazotized. Aromatic diamines are converted to *bis*-diazonium compounds. Diazonium salts are generally used and handled in aqueous solution; they are explosive if isolated and dried.

Coupling reactions. Diazonium ions are mild and selective electrophiles, similar in chemical properties to the nitrosonium ion (NO^+), and they readily and reversibly form covalent bonds to nucleophiles. Reactions with amines lead to diazoamino compounds (triazenes) [reaction (2*a*), where Ar represents the

$$Ar{\equiv}N^+ + RNH_2 \rightleftharpoons ArN{=}N{-}NHR + H^+ \qquad (2a)$$

Triazene

$$ArN{\equiv}N^+ + OH^- \rightleftharpoons [ArN{=}N{-}OH] \xrightarrow{OH^-}$$

Diazohydroxide

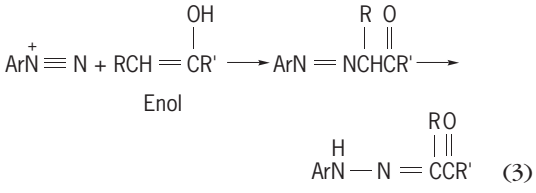
$$ArN{=}N{-}O^- \qquad (2b)$$

Diazotate

aryl group]. With cyanide ion, diazocyanides are

formed. Addition of an azide salt to a diazonium solution at −20°C (−4°F) gives a diazoazide, which cyclizes to the very unstable arylpentazole; the latter decomposes on warming to an aryl azide plus nitrogen. Reaction with hydroxide ion occurs when a diazonium solution is made basic. Since the diazohydroxide is a stronger acid (in the Brönsted sense) than the diazonium ion (Lewis acid), two equivalents of hydroxide are consumed and the product is the diazotate [reaction (*2b*)]. *See* ACID AND BASE; ELECTROPHILIC AND NUCLEOPHILIC REAGENTS.

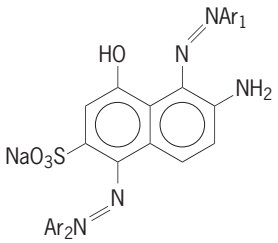
Diazonium salts also react with enols or enolizable carbonyl compounds. The products are hydrazones, which are formed by tautomerization of the azo intermediates [reaction (3)]



(3)

See REACTIVE INTERMEDIATES.

The great importance of diazonium compounds in dye technology lies in the coupling reactions that occur with an activated aromatic ring, such as that in phenols or aromatic amines. Coupling, or electrophilic substitution by ArN_2^+, gives compounds with an arylazo group at the position para or ortho to —OH or —NH_2. Reaction with amines occurs in weak acid solution. With phenols the phenoxide ion is the reactive species, and slightly basic solution is used. These coupling reactions occur concurrently with the equilibria of reactions (*2a*) and (*2b*). A complex balance of reaction rates, acidities, and concentrations is required, but it is possible to prepare compounds in which two different azo groups are introduced, such as the naphthol dye, with the structure below.



See CHEMICAL EQUILIBRIUM.

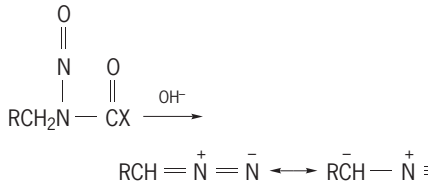
The azo dyes obtained in these coupling reactions are one of the important types of synthetic dyes. The color of the dye can be varied widely by choice of diazonium and coupling components. The coupling reaction lends itself to an important method of applying the dye to fabrics. In this process the coupling reagent, such as a naphtholsulfonic acid, is absorbed onto the fiber, and the coupling reaction is then carried out directly on the fiber by passing the fabric through a bath of the diazonium solution. *See* DYE; DYEING; TEXTILE CHEMISTRY.

Substitution reactions. Diazonium salts are central to the synthesis of a number of aromatic compounds. When a solution of a diazonium salt in water is heated, nitrogen is evolved and the resulting aryl cation is converted to the corresponding phenol. Similarly, aromatic halides and nitriles can be synthesized by treatment of the diazonium salt with an appropriate source of halide and cyanide ions (Sandmeyer reaction). Displacement of nitrogen by hydrogen can be accomplished with hypophosphorous acid. *See* PHENOL.

Diazonium salts are also useful for introducing an aryl group by substitution. Biaryls are formed in a free-radical reaction by heating a neutralized diazonium solution with an aromatic hydrocarbon. Alkenes with an electron-withdrawing group can also be arylated. *See* FREE RADICAL; ORGANIC SYNTHESIS; SUBSTITUTION REACTION.

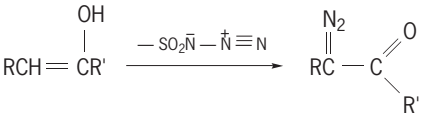
Diazo compounds. In contrast to the aromatic series, diazotization of an alkyl amine gives an extremely unstable and short-lived diazonium ion. Because there is no stabilizing delocalization of positive charge, decomposition occurs to give a carbenium ion and a nitrogen molecule; the reaction products result from the carbocation. However, the diazo group can be retained in an aliphatic or alicyclic molecule if a proton is removed to give the system $R_2C{=}N_2$. These diazo compounds are useful and versatile reagents, and several methods for their preparation are available.

Simple diazoalkanes are prepared by alkaline hydrolysis of a *N*-alkyl-*N*-nitroso amide, sulfonamide, or urea [reaction (*4a*), where X = OR or NH_2]. Another method for introducing a diazo group is direct transfer of the group from a sulfonyl azide to an enolate, resulting in a diazoketone [reaction (*4b*)]; this reaction brings out the similarity of the isoelectronic azido and diazoalkyl groups.



(*4a*)

Diazoalkane

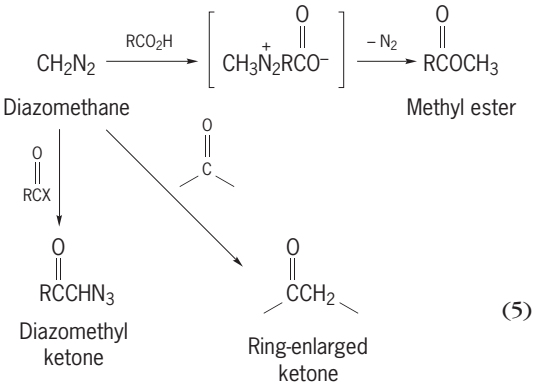


(*4b*)

Diazoketone

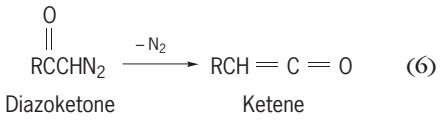
Diazoalkanes are good nucleophiles and react readily with electrophiles such as a proton or carbonyl group. With a carboxylic acid, diazomethane is protonated and the diazonium ion immediately reacts with the anion to give the methyl ester

[reaction (5)]. With a cyclic ketone the major reaction is ring enlargement. Acylation of diazomethane gives the diazomethyl ketone.



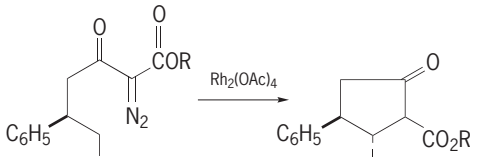
(5)

Diazo compounds and azides on heating or photolysis lose nitrogen to give carbenes and nitrenes, respectively. These are neutral, electron-deficient, highly reactive intermediates that can add to double bonds and also attack C-C or C-H bonds with formation of insertion products. Loss of nitrogen from the acyl derivatives, namely, diazoketones and acyl azides, is accompanied by rearrangement, probably by a concerted pathway rather than a carbene or nitrene, to ketene [reaction (6)] or isocyanate, respectively.



(6)

A very useful reaction of diazoketones or diazoesters involves metal-catalyzed C-C bond formation. Rhodium carboxylates such as dirhodium tetraacetate [$Rh_2(OAc)_4$] are particularly effective catalysts. In these processes the reactivity of the acyl carbene is attenuated by coordination with the metal, and stereospecific cyclizations are possible [reaction (7)].



See COORDINATION CHEMISTRY; REACTIVE INTERMEDIATES.

James A. Moore

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Dichroism

In certain anisotropic materials, the property of having different absorption coefficients for light polarized in different directions.

There are few natural materials which exhibit strong dichroism. One of the first to be discovered was tourmaline. Light transmitted by thin plates of dark forms of tourmaline is almost completely polarized. *See* POLARIZED LIGHT.

In isotropic optical materials the optical density is defined as in Eq. (1), where I_0 is the intensity of the

$$d = \log \frac{I_0}{I} \tag{1}$$

incident light, and I that of the transmitted light. In anisotropic materials that are dichroic, the value of d can vary as a function of the vibration direction of the electric vector of the light wave. Just as the index ellipsoid is used to define the birefringence of a material, a density surface can be used to define the dichroism. *See* CRYSTAL OPTICS.

Compared to the literature on birefringence and optical activity, there has been relatively little material on dichroism. This is partly because of the difficulty in making measurements. The Kramers-Kronig relationship shows that any material whose refractive index is different from unity and varies as a function of wavelength will absorb radiation at some wavelength. From the Kramers-Kronig relationship it is apparent that all optically anisotropic materials should be dichroic. From the values of the refractive index at different wavelengths, the spectral positions and intensity of the absorption can be calculated. In a linear birefringent material the refractive index depends on the polarizing direction or electric vector of the radiation. For each direction of propagation there are two perpendicular vibration directions with different refractive indices. It can be inferred that, at some wavelength, the absorption for light vibrating in these two directions will be different. For transparent materials this wavelength is in the ultraviolet and the absorption is thus difficult to measure. The absorption difference is also frequently so small that it cannot be detected. In other words, the absorption is apparently the same in each direction. Furthermore, the dichroic band may coincide with a region of isotropic absorption.

If the absorption in a dichroic material is different for different linear states of polarization, the material is termed linear dichroic. If it is different for right and left circularly polarized light, it is termed circular dichroic. Similarly, there can be elliptically dichroic crystals. In biaxial crystals there are three different refractive indices corresponding to an electric vector lying along each of the three orthogonal axes of the index ellipsoid. Such a crystal has dichroism which is different for light traveling along each of these three principal axes.

Most dichroic materials exist in the form of relatively thin sheets. Here one is dealing with a section of the dichroic surface. Associated with the sheet of material will be a direction of maximum absorption and one of minimum absorption. The density equation can be rewritten as Eqs. (2), where the densities

$$d_{\parallel} = \log \frac{I_{0\parallel}}{I_{\parallel}} \qquad d_{\perp} = \log \frac{I_{0\perp}}{I_{\perp}} \tag{2}$$

are for light vibrating parallel or perpendicular to the axis of the section of dichroic surface. *See* TRICHROISM. Bruce H. Billings

Dichroism (biology)

In absorbing, anisotropic objects the refractive index and the absorption vary with direction. As with double refraction (birefringence), there are two main directions of anisotropy, which are perpendicular to each other. This phenomenon is called dichroism and may be studied in the polarizing microscope with one polarizer. By rotating the microscope stage the object is brought in different directions to the vibration plane of the polarized light, and the change of light absorption is observed. Study of dichroism allows conclusions as to the submicroscopic fine structure of cells.

Use of stains. In visible light only a few cellular components, such as chloroplasts, show absorption. An absorption can, however, be produced by staining. The dichroic staining of plant fibers is especially simple. The walls of such cells consist of cellulose microfibrils 3–30 nanometers thick. The elongate stain particles of benzidine dyes, for example, congo red, are deposited in an oriented manner in the spaces between the microfibrils and produce an intrinsic dichroism of the fiber: colored for a vibration plane parallel, colorless for a plane perpendicular to the stain particles and fibrils. Therefore, the direction of strongest absorption indicates the course (parallel or helical) of the microfibrils in the fiber. Similar results are obtained by staining with iodine by means of the chlorzinc iodine reagent. Also, some metals, such as silver and gold, give dichroic staining; they are deposited as isotropic crystals. The dichroism is form dichroism (analogous to form birefringence), that is, produced only by the distribution of the crystals. Dichroic staining may also be carried out with animal tissue. Such studies have been made mainly with fibrous protein structures (collagenous fibers, myofibrils, neurofibrils) and with anisotropic nuclei of certain sperms.

Ultraviolet dichroism. In ultraviolet light many cell components show absorption. Ultraviolet dichroism gives direct information as to the orientation of the absorbing molecules or molecular groups in cell structures. The method has been especially helpful for studies of orientation of deoxyribonucleic acid in nuclei and chromosomes. The maximum absorption here is perpendicular to the molecular direction. Dichroism in elongate nuclei of certain sperms shows corresponding behavior, which means that in these objects the deoxyribonucleic acid molecules must be parallel to the long axis. In normal nuclei and in chromosomes there is no or only a very weak ultraviolet dichroism; therefore the deoxyribonucleic acid molecules can have no appreciable orientation.

Lignified plant cell walls show ultraviolet dichroism. It is pure form dichroism, a fact which eliminates the possibility that lignin is in an anisotropic state in the wall. *See* CELL WALLS (PLANT).

By irradiation with ultraviolet light, various compounds of the cell are caused to fluoresce. The fluorescent light is polarized if the object is anisotropic. This phenomenon, called difluorescence, is observable in lignified cell walls, and leads to the same conclusions as to lignin deposition as emerge from dichroism studies. *See* DICHROISM. Fritz Ruch

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Dicotyledons

A large group of flowering plants (angiosperms) that for many years has been considered one of the two main categories of plants, the other being monocotyledons. Dicotyledons have two seedling leaves as opposed to the single one in most monocotyledons. Several deoxyribonucleic acid (DNA) sequence studies subsequently demonstrated that there are two groups of angiosperms, but these correspond not to the number of seed leaves but to the two major pollen types. Thus, the term "dicotyledon" is no longer meaningful because some plants of this type are more closely related to monocotyledons. The group of former dicotyledons, which have pollen with a single aperture, includes magnolia, avocado, black pepper, and pipeworts; they are now termed magnoliids and include monocotyledons. The other category of dicotyledons, those with three (and often more) apertures in their pollen, are called eudicotyledons (true dicotyledons). *See* AVOCADO; EUDICOTYLEDONS; FLOWER; MAGNOLIOPHYTA; MONOCOTYLEDONS; PLANT KINGDOM. Mark W. Chase

Dicranales

A large and diverse order of the true mosses (Bryopsida), containing 13 families. Plants in the Dicranales generally have long, narrow leaves and deeply forked peristome teeth. The plants are erect and simple or merely forked and often woolly because of a dense covering of rhizoids among the leaves. The leaves, inserted in numerous rows, or rarely only two or three are often turned to one side or, when dry, crisped and curled. They are long-tapered, with a long, single costa and short or elongate cells which are sometimes bulging or papillose. The cells at the basal angles are often conspicuously differentiated. The capsules, terminal and generally exserted, may be erect and symmetric or inclined and asymmetric and often furrowed. An operculum is generally differentiated. The peristome (rarely lacking) commonly consists of 16 papillose or vertically pitted-striate teeth. The spores are spherical or rarely tetrahedral. The

calyptra is cucullate or less commonly mitrate. Chromosome numbers 12 and 14 are common, but multiples of 10, 11, 13, and 15 are also known. *See* BRYIDAE; BRYOPHYTA; BRYOPSIDA. Howard Crum

Dicyemida (Rhombozoa)

A phylum (also known as Rhombozoa) comprising approximately 100 described species of microscopic parasites that live in the renal sacs (kidneys) of many species of octopuses, some cuttlefishes, and a few squids.

Life cycle. The dicyemid life cycle is complex and consists of asexual and sexual stages (**Fig. 1**).

Nematogen. When a renal sac of a parasitized cephalopod is opened, or when fluid is withdrawn from the sac by suction, the organisms that emerge are in the asexual nematogen stage. Nematogens are slender and usually about 1 or 2 mm long with an outer layer of up to about 40 ciliated epidermal cells enclosing a large axial cell (**Fig. 2**). The two tiers of ciliated cells at the anterior end form the calotte (Fig. 2). This is inserted into a tubule or pocket in the renal sac, and the short cilia of the calotte cells interweave with microvilli of the epithelium. Nutrients are absorbed from the urine. The free surfaces of most epidermal cells have convoluted ridges that increase the absorptive area, and there are also pinocytotic vesicles where uptake of nutrients may take place.

The axial cell encloses numerous cells called axoblasts (Figs. 2 and **Fig. 3***a*, *b*) which are progeny of one or two cells that entered the axial cell early in the development of the nematogen. The axoblasts develop into daughter nematogens (Figs. 2, 3*a*, *b*), which escape from the parent nematogen and find a vacant site in the renal sac to attach themselves and asexually produce more nematogens.

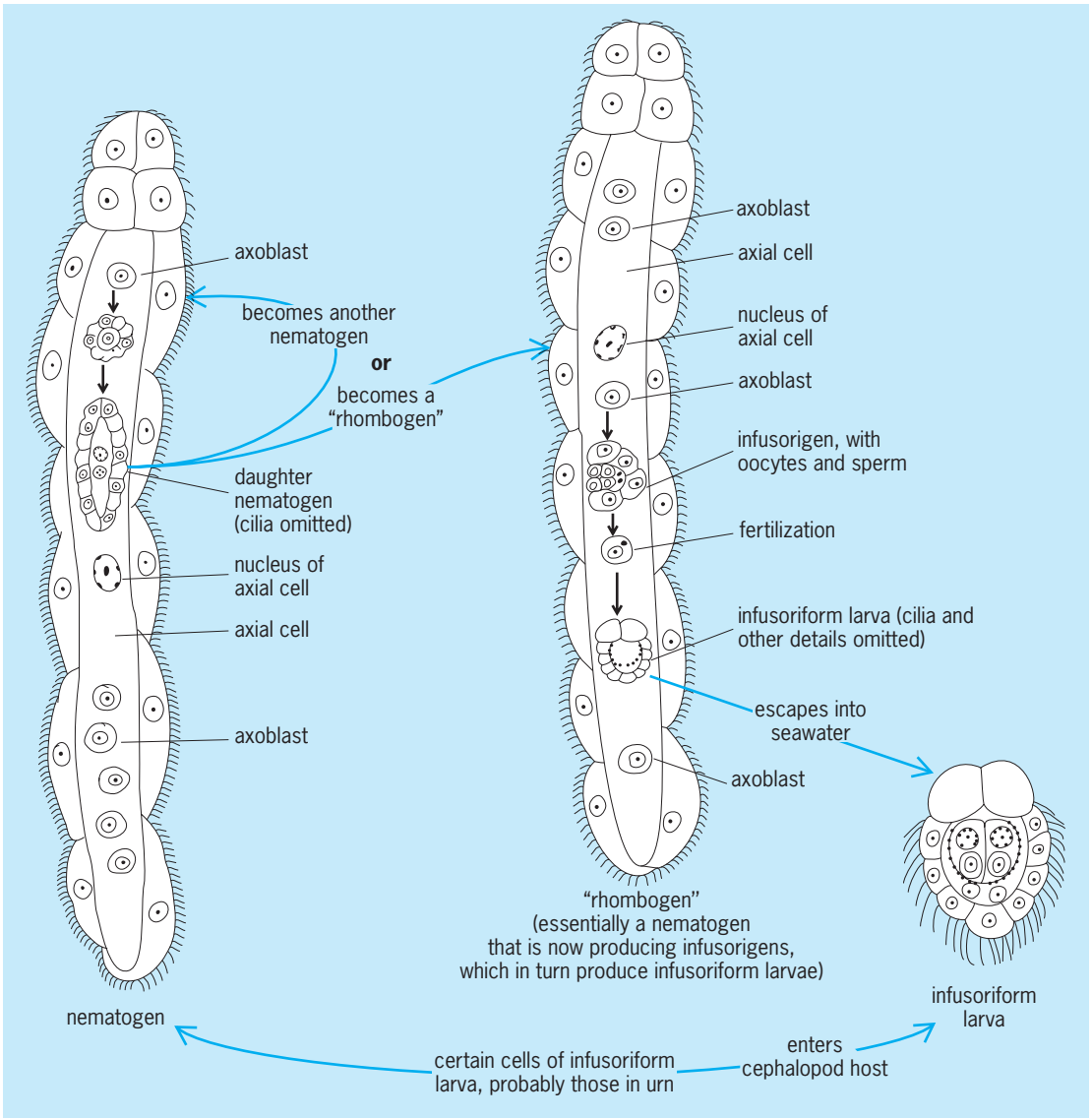

Fig. 1. Diagram of the life cycle of a dicyemid.

Infusorigen. After a nematogen has given rise to many daughter nematogens, this phase of reproduction slows down, and many of the axoblasts begin to develop into infusorigens (Figs. 3*c* and **Fig. 4***a*). These are structurally simple sexual individuals that consist, when fully developed, of a few oocytes and cells that give rise to sperm (which are much smaller than the oocytes and are atypical in that they lack a flagellum). After an oocyte has been entered by a sperm, it undergoes meiosis, casting off polar bodies, and becomes a fertilizable egg, or ovum. The egg nucleus then unites with the sperm nucleus.

Infusoriform larva. A fertilized egg develops into an infusoriform larva (Figs. 3*c* and 4*b*), so called because it resembles an "infusorian" (ciliated protozoan.) In spite of its small size (usually about 50 μm long), the organism at this stage is rather complex. It contains about 37 or 39 cells, two of which are proportionately large and filled mostly by a refringent mass of sodium inositol hexaphosphate. These cells are covered by a thin cap consisting of two cells that lack cilia but may extend posteriorly for a short distance behind the large cells they enclose. The rest of the outer covering of the infusoriform larva consists of ciliated cells. These enclose a few cells, some of which are organized into a structure called the urn, the cavity of which is partly lined by ciliated cells and partly by cells that are not ciliated; certain of the latter contain granules of glycogen. Within the cavity of the urn are four cells, each of which contains one or two nuclei and another complete cell.

A nematogen that has begun to produce infusoriform larvae is often called a rhombogen (hence the alternative name for the phylum), because the larvae have a somewhat rhomboid shape. However, a rhombogen is just an older nematogen, and it does not directly produce infusoriform larvae; these are products of sexual reproduction by an infusorigen. Infusorigens are sometimes regarded as bisexual gonads of a nematogen, but because they originate from axoblast cells they are, like daughter nematogens, essentially separate individuals.

Infusoriform larvae, after escaping from a nematogen in which they have developed, are active swimmers and will survive outside a host for several days. They are believed to represent the infective stage, but it is not known what part of an infusoriform larva becomes transformed into a nematogen. The cells within the urn, each of which contains one or two nuclei as well as another complete nucleated cell, seem like good prospects; the inner cell could conceivably develop into an axial cell and become covered by ciliated epidermal cells derived from the outer portion. This idea, however, is purely speculative.

Classification. There are two families of dicyemids, Dicyemidae and Conocyemidae. Nearly all known species and the two largest genera (*Dicyema* and *Dicyemennea*) belong to the Dicyemidae. In Dicyemidae, the epidermis of the nematogen is cellular and ciliated and the calotte is typically well developed,

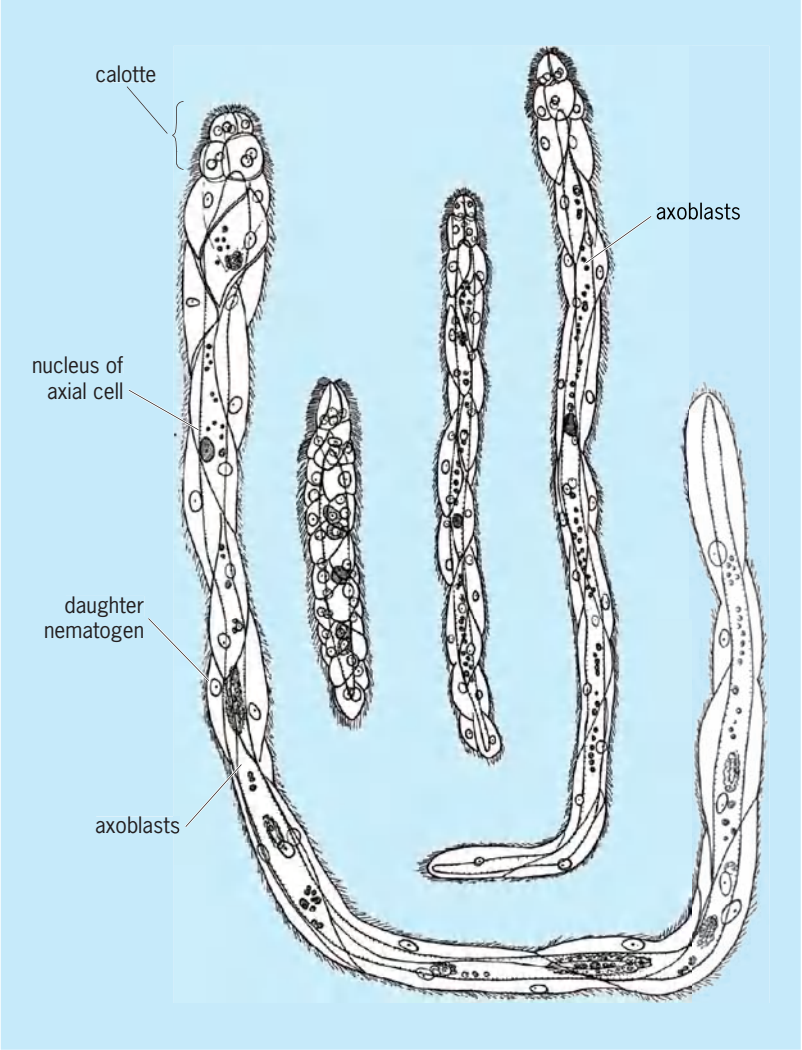

Fig. 2. ***Dicyema*, nematogen stage. The largest specimen contains daughter nematogens. (*From B. McConnaughey, Dicyema sullivani, a new mesozoan from Lower California, J. Parasitol., 35:124, 1949*)**

whereas in Conocyemidae, the epidermis (at least in part) is not divided into cells and not ciliated, and the calotte is not well developed.

Dicyemids were discovered before the middle of the nineteenth century. The earliest serious research on them was published in 1876 by E. van Beneden, who placed them in a category called "Mesozoa" because of his belief that they were intermediate in structure between certain protozoans (especially ciliates) and other "less structurally advanced" invertebrates, such as flatworms. Orthonectida, first reported by Giard in 1877, are very different from dicyemids in structure and life history, but their relative simplicity led to their also being called mesozoans. For a long time, in fact, Mesozoa was a phylum name applied to the two groups. This view should be abandoned, however, because orthonectids and dicyemids are probably not closely related. Either or both groups could conceivably have evolved from free-living ancestral invertebrates or from parasitic groups. Both possibilities seem to be more

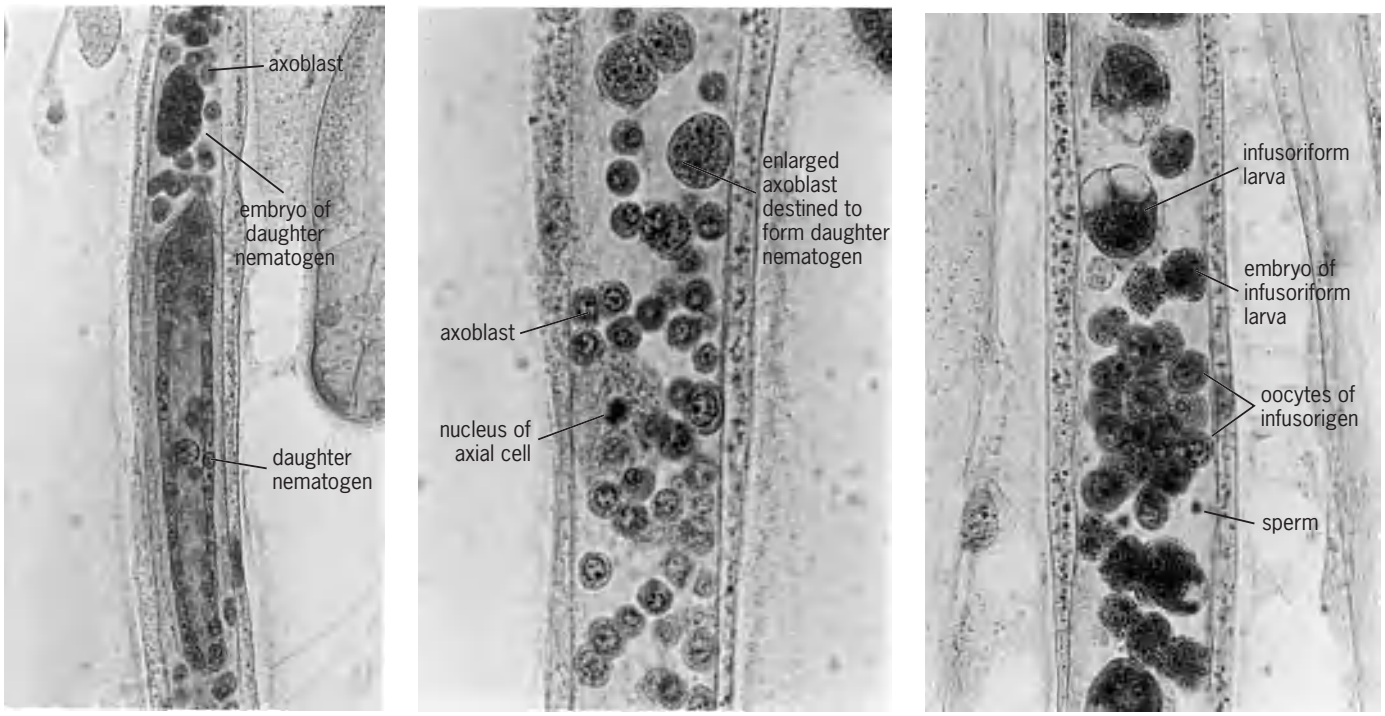

Fig. 3. Photomicrographs of *Dicyemennea*. (*a*) Portion of a nematogen with a daughter nematogen, already ciliated, in its axial cell. (*b*) Portion of an older nematogen whose enlarged axoblasts are destined to develop into infusorigens. (*c*) An infusorigen and stages in the development of an infusoriform larva.

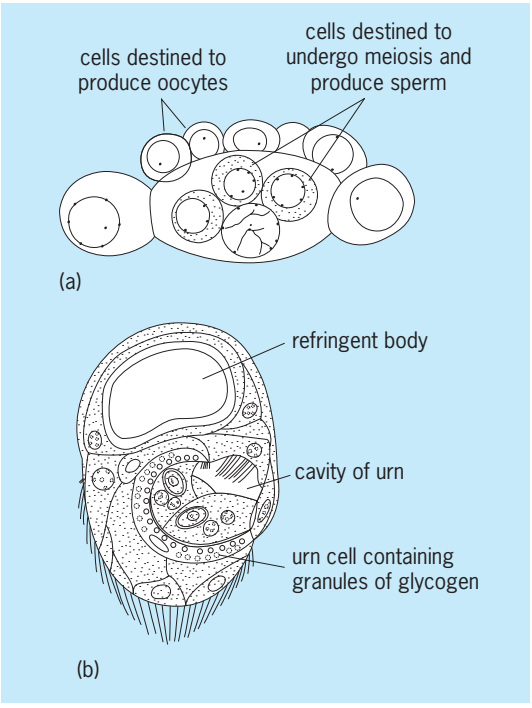

Fig. 4. Early life cycle forms. (*a*) *Dicyema*; young infusorigen with cells destined to produce oocytes and those destined to produce sperm. (*b*) *Dicyemennea*; infusoriform larva, as seen in a sagittal section close to the midline; drawing based on electron micrographs of sections.

popular with specialists than the idea that either dicyemids or orthonectids are derived from protozoa such as ciliates. *See* MESOZOA; ORTHONECTIDA.

Eugene N. Kozloff

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Didymelales

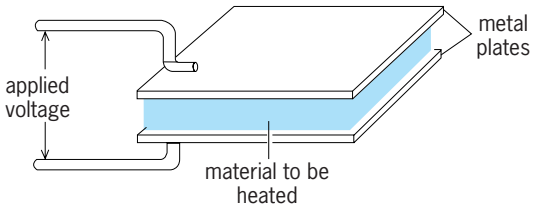
An order of flowering plants, division Magnoliophyta (Angiospermae), in the subclass Hamamelidae of the class Magnoliopsida (dicotyledons). The order contains a single family with but one genus, *Didymeles*, with two species. These dioecious, evergreen trees are restricted to Madagascar. The wood has vessels with scalariform perforations, which is a putatively primitive feature. The leaves are alternate, simple, and entire. The flowers are small; the staminate ones are in open clusters, without perianth, with two stamens, and the pollen grains have three germinal furrows. The pistillate flowers are commonly paired in spikelike clusters; they are hypogynous and sometimes have 1–4 scalelike sepals but no petals. The pistil has but one carpel. This curious genus has often been included in the Hamamelidales; however, the primitive nature of the wood and of the pistil make it anomalous there. *See* HAMAMELIDAE; MAGNOLIOPSIDA; PLANT KINGDOM.

T. M. Barkley

Dielectric heating

The heating of a nominally electrical insulating material due to its own electrical (dielectric) losses, when the material is placed in a varying electrostatic field.

The material to be heated is placed between two electrodes (which act as capacitor plates) and forms the dielectric component of a capacitor, as in the **illustration**. The electrodes are connected to a high-voltage source of 2–90-MHz power that is produced by a high-frequency vacuum-tube oscillator.



Basic assembly for dielectric heating

The resultant heat is generated within the material, and in homogeneous materials is uniform throughout. Dielectric heating is a rapid method of heating and is not limited by the relatively slow rate of heat diffusion present in conventional heating. *See* DIELECTRIC MATERIALS.

Applications. This technique is widely employed industrially for preheating in the molding of plastics, for quick heating of thermosetting glues in cabinet and furniture making, for accelerated jelling and drying of foam rubber, in foundry core baking, and for drying of paper and textile products. Its advantages over conventional methods are the speed and uniformity of heating, which offset the higher equipment costs. Because of the absence of high thermal gradients, an improved end-product quality is usually obtained.

Process. The heating rate obtainable, in watts, is shown by the equation below.

$$P = \frac{1.413AfE^2\kappa' \tan\delta \times 10^{-12}}{d} = \frac{5.56A'fE^2\kappa' \tan\delta \times 10^{-11}}{d'}$$

where P = heating rate, W
A = material area, in.^{2}
A' = material area, m^2
d = thickness of material, or electrode spacing, in.
d' = thickness of material, or electrode spacing, m
f = frequency, Hz
E = voltage across the electrodes, rms
κ' = dielectric constant of material
δ = loss angle (power factor)
$\tan\delta$ = dissipation factor

The factor $\kappa' \tan\delta$ is referred to as the loss factor, and serves to indicate the relative rates of heating for different materials at the frequency to be employed in high-frequency fields of the same intensity. *See* DIELECTRIC MEASUREMENTS.

The loss factor is affected by both temperature and frequency. Materials with a high loss factor heat rapidly, while those with a very low loss factor do not heat at all in high-frequency fields. The high loss factor of water (much higher than most insulators) means that materials which are difficult to heat by dielectric heating when dry are often efficiently heated when moisture is present.

The controllable variables in the process are voltage E and frequency f. Voltage can be raised only within limits determined by corona or dielectric breakdown, generally not exceeding 15,000–20,000 V. Voltage gradient across the work is generally the principal determinant, and this ranges between 1500 and 5000 V per inch (600 and 2000 V per centimeter), depending upon the porosity of the work. *See* CORONA DISCHARGE.

Frequencies used are as high as practicable, so that voltages can be kept within the limits given. Most standard equipment uses frequencies from 5 to 40 MHz, with power outputs up to 125 kW. Power rating decreases for the higher frequencies.

The work electrodes and the material being heated function like a capacitor. They are usually connected as a part of a resonant circuit, which is tuned to the oscillator frequency to obtain maximum transfer of power.

The electrodes are made from nonmagnetic materials, such as aluminum, copper, and stainless steel. The material to be heated is often conveyed through the electrodes on belts made from a low-loss insulating substance, such as coated glass cloth. *See* ELECTRODE.

Radiation shielding. These are certain frequencies, such as 13.56, 27.12, and 40.68 MHz, specifically allocated by international agreement for industrial use. In addition, to prevent interference with authorized radio services, such as established radio communication channels, industrial radio-frequency (rf) heating installations operating above 10,000 Hz must be properly shielded.

The usual method is to enclose completely all rf circuits and electrodes in a metal cabinet, using wire-mesh shielding over ventilation openings and good electrical contacting or overlapping surfaces on access doors. Where inlet and exit openings must exist, as in conveyorized installations, a metallic shield in the form of a vestibule or throat is used. The length and area of the shield are designed to attenuate rf radiation from the electrode area.

George F. Bobart

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Dielectric materials

Materials which are electrical insulators or in which an electric field can be sustained with a minimal dissipation of power. A solid is a dielectric if its valence band is full and is separated from the conduction band by at least 3 eV.

Dielectrics are employed as insulation for wires, cables, and electrical equipment, as polarizable media for capacitors, in apparatus used for the propagation or reflection of electromagnetic waves, and for a variety of artifacts, such as rectifiers and semiconductor devices, piezoelectric transducers, dielectric amplifiers, and memory elements.

The ideal dielectric material does not exhibit electrical conductivity when an electric field is applied. The term dielectric, though it may be used for all phases of matter, is usually applied to solids and liquids. In practice, all dielectrics do have some conductivity, which generally increases with increase in temperature and applied field. For a good dielectric, such as polytetrafluoroethylene, the low-field direct-current conductivity at room temperature may be lower than $10^{-16}\ \Omega^{-1} \cdot m^{-1}$, whereas the corresponding figure for some specimens of plasticized poly(vinyl chloride) may be as high as $10^{-4}\ \Omega^{-1} \cdot m^{-1}$, similar to that of the low-conductivity semiconductors.

Breakdown. If the applied field is increased to some critical magnitude, the material abruptly becomes conducting, a large current flows (often accompanied by a visible spark), and local destruction occurs to an extent dependent upon the amount of energy which the source supplies to the low-conductivity path. This critical field depends on the geometry of the specimen, the shape and material of the electrodes, the nature of the medium surrounding the dielectric, the time variation of the applied field, and other factors. Temperature instability can occur because of the heat generated through conductivity or dielectric losses, causing thermal breakdown.

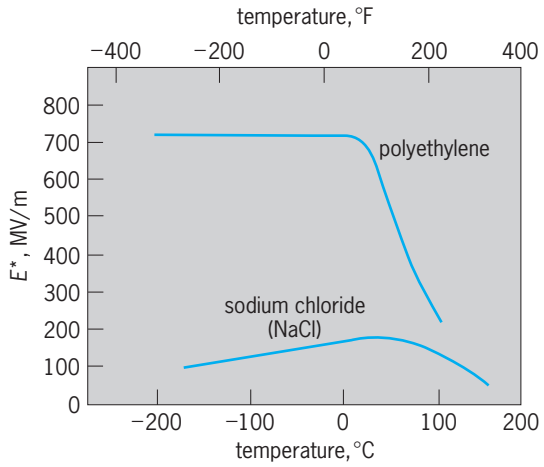
Solids may contain gaseous voids, or such voids may appear and grow in service. Since the gas will have a lower electric strength than the dielectric medium, gas discharge is likely in the voids, with accompanying destruction of the adjacent material and field intensification nearby. Thus the void grows and initiates a tree, and ultimately breakdown takes place. The same consequence may follow from the growth of water trees caused by the absorption of moisture into the bulk of the dielectric. Surface tracking is a common cause of breakdown, when a conducting path appears and elongates on the surface between the dielectric and the ambient medium, due to destruction of the material by localized discharges. After long exposure to the atmosphere, chemical degradation of solid dielectrics can occur, and so breakdown can be encouraged. In liquids, particles can be drawn into the high-field region, causing partial bridging of the high-potential gap, and so precipitate breakdown.

Thus, breakdown can be brought about by a variety of different causes, sometimes by a number of them acting simultaneously. Nevertheless, under carefully specified and controlled experimental conditions, it is possible to measure a critical field which is dependent only on the inherent insulating properties of the material itself in those conditions. This field is called the intrinsic electric strength of the dielectric. *See* ELECTRICAL BREAKDOWN.

Solids. The theory of dielectric breakdown was developed by the use of alkali halide crystals as a model, since their simple cubic lattice structure aids theoretical analysis. Although the gap between the valence and conduction bands is substantial (say 7 eV), crystal imperfections will result in isolated levels with a ground state (say 1 eV) and excited states (perhaps 0.1 eV) below the conduction band. Thus, thermal ionization ensures that there are always some electrons in the conduction band available to be accelerated by an applied field. The energy they gain is given up to the lattice by collisional excitation of vibrational modes. For the intrinsic breakdown field E^*, the power absorbed by the lattice is just less than the power gained from the field by electrons in a certain critical energy range; these electrons then accelerate up to ionization energy, and electron multiplication by an avalanche process ensues over a sufficient number of generations for local material destruction to occur. Since the electron scattering by the lattice increases with the lattice's deviation from periodicity, it would be expected that the intrinsic strength would increase with temperature, which is confirmed experimentally. *See* BAND THEORY OF SOLIDS; IONIC CRYSTALS.

However, if the temperature is increased further, the number of electrons in excited levels of the isolated states and in the conduction band ultimately becomes so high that electron-electron collisions predominate, and since the transfer of energy to the lattice is rather slow, the electron temperature rises above the lattice temperature. Provided the field is not too high, equilibrium can be attained, but there exists a critical field beyond which this is not so, and this defines the intrinsic strength. Analysis shows that this decreases with increasing temperature, and again, this is confirmed experimentally. Alkali halide crystals therefore exhibit a maximum electric strength, usually near room temperature (see **illus.**). Breakdown caused by these intrinsic processes is very fast, of the order of nanoseconds or less.

For materials with a disordered structure, in the low-temperature region the electric strength would be expected to be more and the slope less, possibly effectively zero. The transition to the high-temperature region tends to occur at a lower temperature, as the number of ionizable centers is large. The decline in strength is then quite sharp. This behavior is typical for polymers (see illus.). Although this behavior is in accordance with the theory outlined above, other explanations have been put forward, and the mechanism of intrinsic breakdown cannot be considered to be definitely



Variation of electric strength with temperature for sodium chloride and polyethylene.

established, particularly in the high-temperature region.

Another mechanism causing breakdown is thermal runaway caused by Joule heating. Because the material has some conductivity, and dielectric loss if the applied voltage is alternating, heat is generated. As the temperature rises, so does the conductivity, and if the generated heat cannot reach equilibrium with the heat loss at a sufficiently low temperature, then the material is destroyed. This is a much slower process than the electronic processes discussed above, and for most practical cases, the time for it to occur would rarely be less than about 1 millisecond, and often very much longer. Thermal breakdown often occurs in polymers, especially when the ambient is significantly above room temperature. *See* JOULE'S LAW.

Liquids. Less is known about the mechanism of liquid than of solid breakdown. Nevertheless, the development of very sophisticated photographic techniques has made it possible to study the critical processes which take place a few nanoseconds before breakdown occurs. These processes are complicated, and not necessarily the same for all liquids and conditions. In general, light detectable by image-intensifier techniques appears in the gap up to 1 microsecond or as little as 20 nanoseconds before breakdown, depending on electrode shape. This stage is probably associated with streamer formation and the growth of filamentary space-charge channels through the liquid. It is followed by the appearance of low-density disturbances, in some cases perhaps bubbles or cavities, near the cathode. These grow until the gap is bridged and breakdown occurs. The dielectric strength tends to increase with pressure and diminish with increasing temperature. Generally, liquids have a lower strength than solids; for example, to compare two good insulants at room temperature, polyethylene has a strength of 500–700 MV/m while transformer oil has a strength of 110–130 MV/m.

In large power devices, dielectric liquids provide not only insulation but also circulatory cooling. They also perform the important function of filling voids in solid insulation, for example, cables or high-voltage capacitors. It is an advantage for liquids used for impregnation to have a high E^* and a high ϵ_r' (real part of the complex relative permittivity), because this reduces the field in the filled cavity. Formerly, the liquids used for this purpose were the chlorinated biphenyls, which have these advantages and additionally have a low viscosity and are fire resistant. Unfortunately, they were found to be dangerous biochemically. However, satisfactory performance has since been achieved by the use of various esters. *See* POLYCHLORINATED BIPHENYLS.

Industrial dielectric materials. Many of the traditional materials are still in common use, and they compete well in some applications with newer materials regarding their electrical and mechanical properties, reliability, and cost. For example, oil-impregnated paper is still used for high-voltage cables, usually paper with values of ϵ_r' about 3–3.5 and tan δ (the loss tangent; the negative of the ratio of the imaginary part to the real part of the complex relative permittivity) about 2×10^{-3} at room temperature, rising to about 3×10^{-3} at 100°C (212°F); various types of pressboard and mica, often as components of composite materials, are also in use. Elastomers and press-molded resins are also of considerable industrial significance. However, synthetic polymers such as polyethylene, polypropylene, polystyrene, polytetrafluoroethylene, polyvinyl chloride, poly(methyl methacrylate), polyamide, and polyimide have become important, as has polycarbonate because it can be fabricated into very thin films.

Properties of polymers. Generally, these have crystalline and amorphous regions, increasing crystallinity causing increased density, hardness, and resistance to chemical attack, but often producing brittleness. Many commercial plastics are amorphous copolymers, and often additives are incorporated in polymers to achieve certain characteristics or to improve their workability. Additives include inorganic fillers, antioxidants, stabilizers, and pigments.

Polyethylene. Polyethylene (C_nH_{2n+2}) is the most common and most investigated polymer. It is relatively inexpensive and is easily worked. Polyethylene readily oxidizes, and so its working temperature is normally not above 70°C (158°F), although it may be worked at higher temperatures (say up to 90°C or 194°F) in suitable circumstances; it normally contains antioxidants. The molecular chains form planar zigzags, each chain being roughly perpendicular to those faces of the crystal plate containing it. Since the chain length is generally much longer than the distance between the faces of the plate (about 10 nanometers), the chain is generally folded back when it reaches the plate surface, although it can pass out into amorphous regions or other crystal plates. In its low-density form, polyethylene has a few branches along the molecular chain and is

usually 50–60% crystalline. *See* POLYOLEFIN RESINS.

High-density polyethylene has very few branches, is about 90% crystalline, and generally has better mechanical properties than those of the low-density form; its electrical properties are similar. Polyethylene is used for communications cables, and cross-linked polyethylene is used for high-voltage power cables. Here, voids may lead to treeing, and in a very moist environment a destructive phenomenon, water treeing, may occur. Polyethylene is also used for a multitude of small, inexpensive components. At room temperature, E^* is 500–700 MV/m; and over a wide frequency range, for commercial polyethylene ϵ'_r is 2.3 and tan δ is usually about 2.5×10^{-4} but may be more or less by a factor of 2.

Poly(ethylene terephthalate). Poly(ethylene terephthalate) is a useful material which has good resistance to acids and alkalies provided that they are dilute. It is insoluble in many common solvents, although it does dissolve in aromatic and chlorinated hydrocarbons, especially when warm. It softens at about 250°C (480°F).

Polypropylene. Polypropylene is similar in structure to polyethylene, but every other carbon atom has one of its hydrogen atoms replaced by a CH_3 group. Though electrically similar to polyethylene, polypropylene can be made in thinner films, say 5 micrometers as against about 25 μm for polyethylene. These films replace paper for impregnated capacitors, with reduced loss.

Polystyrene. Polystyrene is brittle at room temperature, becomes soft at 80°C (176°F), and is often modified by copolymerization. Traditionally, it is used in film form for capacitors, and it remains competitive for this application. Polystyrene is also used for coaxial-cable insulation, but in wound strip or bead form, since the solid is not very flexible. With very low water absorption, tan δ can be around 10^{-4} up to 1 MHz. For commercial polystyrene, ϵ'_r is about 2.6 and E^* about 400–600 MV/m. *See* POLYSTYRENE RESIN.

Polytetrafluoroethylene. Polytetrafluoroethylene is similar to polyethylene, with all the hydrogen atoms replaced by fluorine atoms. Highly crystalline (about 95%), it resists twisting and bending and is mechanically tough, with a very low coefficient of friction, 0.06. It is highly resistant to chemical attack and is useful in hostile environmental conditions. The loss is very low, with tan δ less than 2.5×10^{-4} at frequencies up to 1 GHz, and tan δ less than 10^{-4} at frequencies below 1 MHz; E^* is in the range 400–500 MV/m, and ϵ'_r is almost constant at 2.1 over a wide frequency range. *See* POLYFLUOROOLEFIN RESINS.

Poly(vinyl chloride). Poly(vinyl chloride) has a structure similar to that of polyethylene, but every other carbon atom has one of its hydrogen atoms replaced by a chlorine atom. It is only about 10% crystalline and is compatible with a large number of other polymers. It is easily decomposed by heat and is not suitable for continuous use above 70°C (158°F). Poly(vinyl chloride) is not well characterized, and depending on admixtures, may be rigid or flexible. Though suitable compositions may be molded, its main electrical use is for coating wires. Water absorption, especially by rigid poly(vinyl chloride), is high, but absorption may be kept down by suitable fillers. Its ϵ'_r is usually between 3.5 and 12, and tan δ may be as large as 0.2. It does not withstand high electrical stresses. *See* POLYVINYL RESINS.

Poly(methyl methacrylate). Poly(methyl methacrylate) is a clear solid, hard but easily scratched. Its dielectric properties are only moderate, and its use is restricted to undemanding conditions at normal ambient temperatures, usually in a molded form, where appearance is important. Its ϵ'_r is about 3.6, with tan δ about 0.05. *See* POLYACRYLATE RESIN.

Polyamide. Polyamide may be in linear form, known as nylon, or in aromatic form. Neither of these is of much value as a dielectric, but nylons, because of their toughness and resistance to solvents, are used to form coatings to protect insulation. The aromatic form can also make yarns, but it is most often used in impregnated board form in which it gives low-voltage insulation capable of being used continuously at temperatures of 150°C (302°F), or even higher. Its ϵ'_r is about 3.4, with tan δ about 0.03. *See* POLYAMIDE RESINS.

Polyimide. Polyimide is related to aromatic polyamide, but with additional aromatic groups. Expensive to produce, it can be made in films down to 25 μm thick, and in varnish form for wire coating. It is very tough, does not burn but chars above 800°C (1500°F), and remains flexible at liquid helium temperatures; it can be used continuously up to 250°C (482°F). Polyimide resists organic solvents and acids but not strong alkalies, has a high water absorption rate, and is subject to surface tracking. Its ϵ'_r is about 3.4, with a tan δ of 0.005. *See* HETEROCYCLIC POLYMER.

Resins. The resins are important members of the family of cross-linked polymers. Epoxy resins have a high mechanical strength, absorb very little water, and bond easily to most materials but not to polyethylene. They are used for bonding and encapsulation, and their properties can be modified by the curing process used and by the use of fillers and hardeners. The ϵ'_r of epoxy resins is about 3.6–4.0. and tan δ is 0.01–0.03 at room temperature at frequencies below 1 MHz. *See* ELECTRICAL INSULATION; PERMITTIVITY; POLYETHER RESINS; POLYMER. James H. Calderwood

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Dielectric measurements

Measurements of the dielectric properties of a material, which are characterized by its complex relative permittivity ϵ_r. For all materials except ferroelectrics, this quantity does not depend on applied field: the general behavior is linear, and so voltage of any convenient magnitude can be used for measurement. *See* FERROELECTRICS; PERMITTIVITY.

Bridge methods. The most commonly used apparatus for measuring ϵ_r is the alternating-current (ac) bridge. These bridges are readily available in the operating range 10–10^6 Hz, and sometimes outside it; ultralow-frequency bridges can go as low as 10^{-3} Hz. Most specimen holders for solids are essentially parallel-plate capacitors with the specimen filling all the space between the circular plates; for liquids, a test cell with cylindrical electrodes is usually employed. Guard electrodes are incorporated to avoid errors due to field fringing and surface conduction, although at the higher frequencies guard electrodes as well as the lead impedances cause difficulties, and it is usual to employ specially designed two-terminal specimen holders.

Derivation of dielectric properties. The measured impedance may be expressed in terms of an equivalent capacitance C_p and resistance R_p in parallel. From these, at an angular frequency ω, the real part of the complex relative permittivity, ϵ_r', is given by Eq. (1); the negative of the imaginary part, ϵ_r'', is given by Eq. (2); and the loss angle δ is given by Eq. (3), where C_0 is the capacitance of the empty specimen holder.

$$\epsilon_r' = \frac{C_p}{C_0} \tag{1}$$

$$\epsilon_r'' = \frac{1}{\omega C_0 R_p} \tag{2}$$

$$\tan\delta \equiv \frac{\epsilon_r''}{\epsilon_r'} = \frac{1}{\omega C_p R_p} \tag{3}$$

Alternatively, the measured impedance may be expressed in terms of an equivalent capacitance C_s and resistance R_s connected in series. In terms of these, ϵ_r', ϵ_r'', and tan δ are given by Eqs. (4), (5), and (6).

$$\epsilon_r' = \frac{C_s}{C_0\,(1+\tan^2\,\delta)} \tag{4}$$

$$\epsilon_r'' = \frac{C_s\tan\delta}{C_0\,(1+\tan^2\,\delta)} \tag{5}$$

$$\tan\delta = \omega C_s R_s \tag{6}$$

If the material contains free charge carriers resulting in an ohmic conductivity σ, then this causes an apparent ϵ_r'' given by Eq. (7). This quantity has to be

$$\epsilon_r'' = \frac{\sigma}{\epsilon_0\omega} \tag{7}$$

subtracted from the measured ϵ_r'' to give the value of ϵ_r'' attributable to dielectric loss processes.

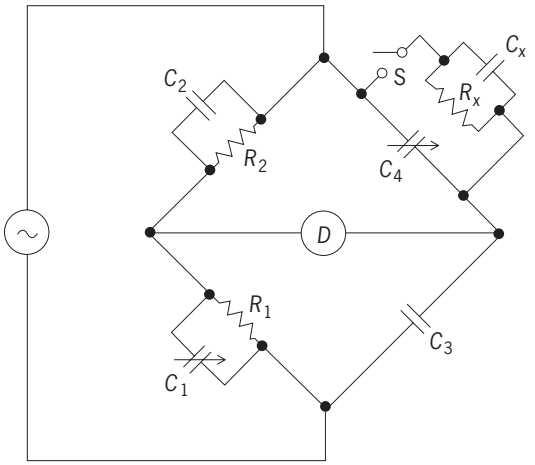

Fig. 1. High-frequency form of the Schering bridge. *D* = detector; *S* = switch.

Schering bridge. The bridges most commonly used are of the Wheatstone type, the most versatile for dielectric measurements being the Schering bridge. This is commonly used in its conjugate form arranged for substitution measurement (**Fig. 1**). The dielectric specimen is represented by the elements C_x and R_x in parallel. The bridge is first balanced (that is, adjusted until the voltage across the detector is zero) with a switch between the specimen and the bridge circuit open, and the values of adjustable capacitors C_1 and C_4 on opposite arms of the bridge noted. The switch is then closed so that the specimen is in parallel with C_4, and balance is regained by changing C_1 and C_4 by amounts ΔC_1 and ΔC_4, respectively. It may be shown that, provided tan δ_x is less than 10^{-2}, the capacitance and loss angle of the specimen are given to a good approximation by Eqs. (8) and (9). By the use of standard Wagner grounding

$$C_x = \Delta C_4 \tag{8}$$

$$\tan\delta_x = \omega R_1 C_4 \frac{\Delta C_1}{\Delta C_4} \tag{9}$$

techniques and grounded screens around all bridge components, high-precision measurements of ϵ_r' and tan δ may be made up to a frequency 10^6 Hz. *See* BRIDGE CIRCUIT; WHEATSTONE BRIDGE.

The Schering bridge may also be used in a different form for high-voltage measurements at 60 Hz (**Fig. 2**). R_3 and C_1 are adjusted until balance is achieved, and the capacitance and loss angle of the specimen are then given by Eqs. (10) and (11). Here, R_1 and C_1

$$C_x = \frac{C_2 R_1}{R_3\,[1+(\omega C_1 R_1)^2]} \tag{10}$$

$$\tan\delta_x = \omega C_1 R_1 \tag{11}$$

are a parallel resistance and capacitance on the arm of the bridge opposite the specimen, and C_2 and R_3 are a capacitance and a resistance on the arms adjacent to the specimen. Because the test voltage is

high, the high-potential components are surrounded by a grounded cage. The voltage appearing outside the cage is small; for example, if the test voltage is 100 kV and C_2 is 100 picofarads, then the potential difference across the variable bridge arms is about 1 V. Because the frequency is also low, R_3 and C_1 may be of the decade-box type. However, a protective device must be incorporated across the low-voltage bridge arms in case the specimen should break down and a high voltage appear outside the cage and cause damage to the bridge elements and create a hazard for the operator.

Other bridge types. The are several other four-armed types of bridge, such as the bridged-T network, which has the advantage that the source and detector have a common terminal which may be grounded. Operating on a different principle, bridges with inductively coupled ratio arms compare the impedance to be measured directly with standard components. The balance condition of such a transformer bridge (**Fig. 3**) is given by Eq. (12).

$$\frac{Z_1}{Z_2} = \frac{N_2 N_4}{N_1 N_3} \tag{12}$$

The advantages of this system are that both the source and detector may be grounded at one point and that three-terminal devices cause no difficulties, as the guard ring is connected to point G. The stray admittances associated with the guard ring then do not affect the balance.

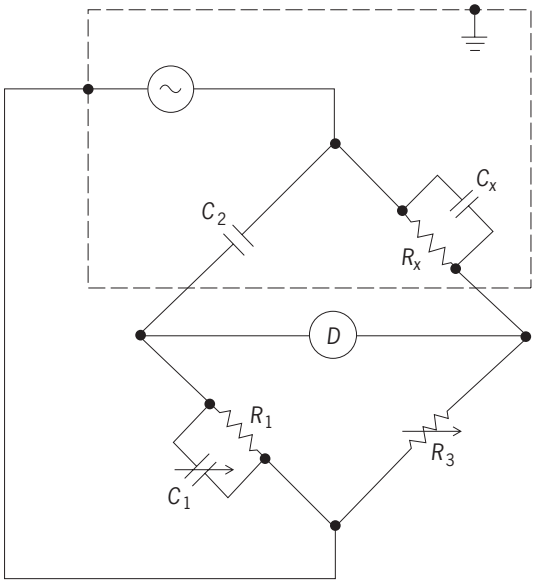

Fig. 2. High-voltage form of the Schering bridge.

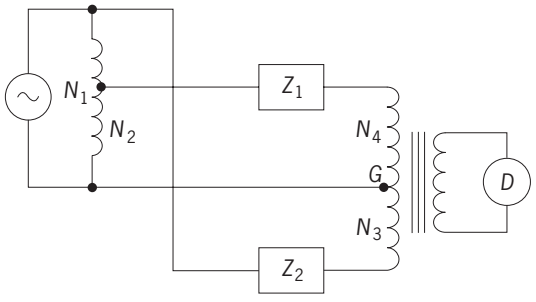

Fig. 3. Transformer bridge.

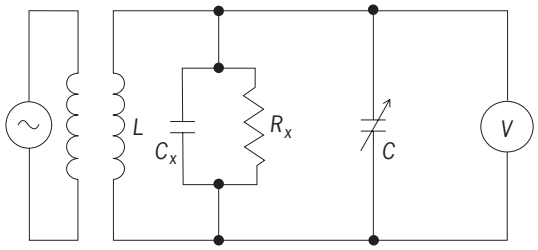

Fig. 4. Hartshorn and Ward basic circuit.

Resonance methods. Resonance methods, useful for frequencies greater than 1 MHz, involve the injection of voltage or current by one of several methods into an *LC* (inductance-capacitance) resonant circuit. The classical circuit arrangement is that of L. Hartshorn and W. H. Ward (**Fig. 4**). Two circular metal plates are in contact with the specimen in the shape of a disk of the same radius to form a capacitor with plate separation x, represented by elements C_x and R_x in parallel, the distance between the plates being controllable by a micrometer screw; a variable capacitor C is in parallel with the specimen. The circuit is energized by the employment of a loosely coupled oscillator, the frequency of which is adjusted for resonance as indicated by the maximum voltage V_i registered by a voltmeter V. The specimen is now removed, and resonance is restored by reducing x by an amount δx by the micrometer screw, when the capacitance between the plates is C_0, and the new maximum voltage V_0 is noted. The auxiliary variable C is now used to measure the change in capacitance δC necessary to cause a shift between the two half-power points on either side of resonance. It is evident that ϵ_r' is given by Eq. (13), and it can be shown that tan δ is given by Eq. (14).

$$\epsilon_r' = \frac{x}{x - \delta x} \tag{13}$$

$$\tan \delta = \frac{[(V_0/V_i) - 1]\,\delta C}{2C_0} \tag{14}$$

See RADIO-FREQUENCY IMPEDANCE MEASUREMENTS; RESONANCE (ALTERNATING-CURRENT CIRCUITS).

Measurements over a range of frequencies may be made by using coils with different inductance values, but ultimately the inductance required becomes impracticably small, and in the range 10^8–10^9 Hz reentrant cavities are often used. These are hybrid devices in which the plates holding the specimen still form a lumped capacitor, but the inductance and capacitances are distributed along a coaxial line. At higher frequencies, the wavelength is comparable to the dimensions of the apparatus, and transmission methods in coaxial lines and waveguides can be used.

Resonance methods can also be used at microwave frequencies, for example, by using resonant cavities

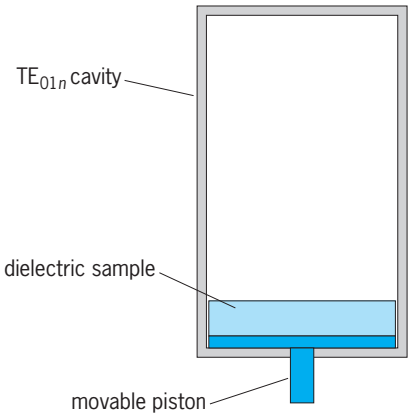

Fig. 5. TE_{01n} resonator. The resonator can be brought into resonance by moving the piston or by changing the frequency.

where the permittivity of the sample can be deduced from the change in the resonant frequency or length and Q of the cavity. TE_{01n} cavities (**Fig. 5**) need disk-shaped samples and TM_{01n} cavities need thin cylindrical samples. Resonance methods are usually limited to measurements on low-loss materials but can measure higher-loss materials by employing perturbation techniques. *See* CAVITY RESONATOR; Q (ELECTRICITY).

Open resonators measure low-loss materials. In a typical arrangement using plane and curved metal mirrors (**Fig. 6**), a Gaussian beam is generated within the resonator and the sample is placed at the beam waist on plane mirror. The microwaves are usually introduced into the cavity by two holes in the center of the curved mirror. Alternative arrangements involve two curved mirrors. Because of the small beam size, open resonators are used for frequencies above 10 GHz.

Other more specialized techniques include dielectric resonators, split-post dielectric resonators, and stripline resonators. *See* MICROWAVE.

Transmission and reflection methods. Coaxial lines are used in the frequency range 300 MHz–3 GHz, and waveguides in the range 3–30 GHz. The transmission characteristics are determined by the complex permittivity of the material filling the line or guide. Many different measurement techniques have been devised, but all derive values of the complex relative permittivity ϵ_r from its relationship to the complex propagation factor γ, that is, Eq. (15), where λ_0 is the

$$\epsilon_r = -\gamma^2 \left(\frac{\lambda_0}{2\pi}\right)^2 \tag{15}$$

free-space wavelength for the frequency employed. The propagation factor γ is obtained by measuring its real and imaginary components α and β, where α gives the attenuation, and β the wavelength λ in the material. For a traveling wave in a coaxial line, it may be shown that ϵ_r' and ϵ_r'' are given by Eqs. (16) and (17).

$$\epsilon_r' = \left(1 - \frac{\alpha^2}{\beta^2}\right)\left(\frac{\lambda_0}{\lambda}\right)^2 \tag{16}$$

$$\epsilon_r'' = \frac{2\alpha}{\beta}\left(\frac{\lambda_0}{\lambda}\right)^2 \tag{17}$$

For transmission in a waveguide limited to the single transverse TE_{01} mode, these equations are somewhat modified, since ϵ_r is given by Eq. (18), where γ_g

$$\epsilon_r = \frac{\left(\frac{1}{\lambda_c}\right)^2 + \left(\frac{\gamma_g}{2\pi}\right)^2}{\left(\frac{1}{\lambda_c}\right) + \left(\frac{1}{\lambda_{0g}}\right)^2} \tag{18}$$

is the propagation factor of this mode in the material-filled guide, λ_{0g} is the mode wavelength in vacuum, and λ_c is the cutoff wavelength of the guide. *See* TRANSMISSION LINES; WAVEGUIDE.

In practice, traveling waves are rarely used as the basis of measurement, except for high-loss materials. Usually, reflections from terminations set up standing waves, the amplitude of which in the case of a liquid-filled line can be measured by a suitable probe. The ratios of the field magnitudes at adjacent maxima, and the distance between them, give the information required. For solid, ϵ_r can be determined by probe measurements in an air-filled section of line between the signal source and a dielectric-filled section. Other methods, more often used for waveguide measurements, involve the use of a movable short-circuiting piston. *See* ATTENUATION (ELECTRICITY); MICROWAVE IMPEDANCE MEASUREMENT; WAVELENGTH MEASUREMENT.

Dielectric properties can also be measured with open-ended waveguide or coaxial probes placed directly on the surface of the material (**Fig. 7**). Coaxial probes are more convenient than waveguides because of their greater frequency range and generally smaller size. These devices are widely applicable to noninvasive dielectric measurements, including measurements on solids, liquids, and biological tissue. The traveling wave propagates along the coaxial line to its end, where fringing fields are launched

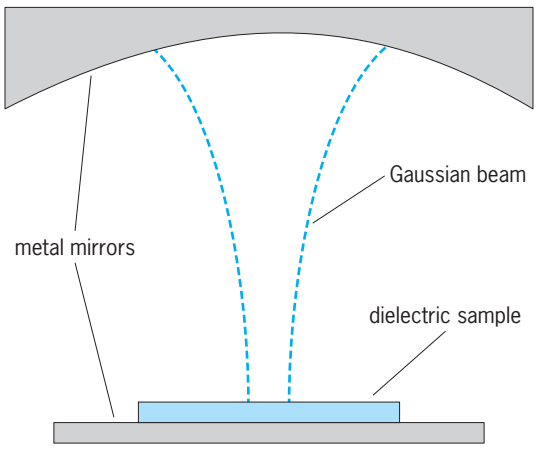

Fig. 6. Open resonator. The microwave radiation is usually introduced via waveguides through holes in the center of the curved mirror.

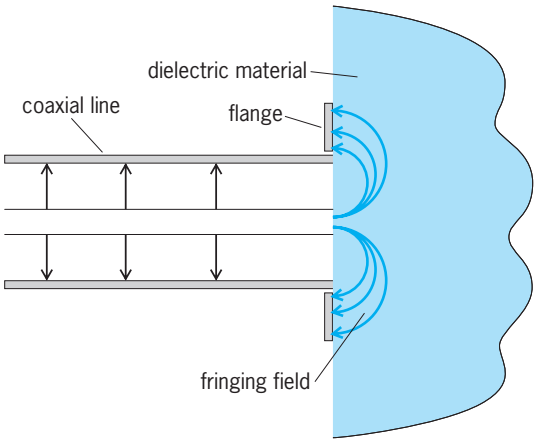

Fig. 7. Coaxial probe.

into the material. The reflection coefficient of the probe is a function of the dielectric properties of the material. The probes are often available with tested software to relate the measured reflection to the material properties. The permittivity of the material can be related to the reflection coefficient by modal analysis of the fields in the coaxial line and an analysis of the fields in the dielectric where the probe is treated as an antenna. In some cases at lower frequencies a capacitive model can be used for the fringing fields. *See* ANTENNA (ELECTROMAGNETISM).

At higher frequencies, free-space or optical-methods can also be used, in which the microwave beam is launched by a horn and can be collected by another horn. The sample is placed in the beam between the horns and its dielectric properties can be deduced from its effect on the beam. This method can be used in transmission or reflection. If corrugated horns are used in conjunction with lenses, a well-shaped Gaussian beam can be produced and this leads to more accurate measurements. Both this method and the open resonator method described above can investigate anisoptropic dielectrics.

Submillimeter measurements. Dielectric measurements are difficult to carry out in the frequency range 30–300 GHz, for which λ_0 is in the range 1 cm–1 mm, but for λ_0 less than 1 mm, methods related to infrared spectroscopy are used. Broadband continuous spectra result from Fourier transform spectroscopy, which in its simplest form is equivalent to normal infrared spectroscopy, with the specimen in one of the two passive arms of the interferometer, between the beam divider and either the source or the detector. In the more sophisticated dispersive Fourier transform spectroscopy, which is more difficult experimentally but offers some procedural and data-handling advantages, the specimen is in one of the active arms, that is, between the beam divider and either mirror (**Fig. 8**). Broadband spectroscopy is only one possible measurement technique; discrete-point spectra also may be obtained by the use of a Mach-Zehnder interferometer and a laser source. By using interferometric techniques, the frequency range can be extended up to about 5 THz. Above that frequency is the region of resonant processes, the domain of the established methods of infrared and optical measurements. *See* INFRARED SPECTROSCOPY; INTERFEROMETRY; SPECTROSCOPY; SUBMILLIMETER-WAVE TECHNOLOGY.

Time-domain methods. If a constant direct-current (dc) voltage is suddenly applied to a dielectric specimen, in principle the charging current is related through the Fourier integral transformation to the steady-state ac current which would flow if the applied voltage were sinusoidal at any particular frequency. If the dc voltage is suddenly removed, a similar relationship holds between the discharge current and ac current. Thus the variation of complex permittivity with frequency can in principle be derived from a transient signal in the time domain. Because of various limitations, the method is not capable of giving results of an accuracy at all frequencies comparable to those obtainable from a single frequency measurement. Nevertheless, with the aid of computer analysis, the response over a large frequency range can be obtained much more quickly than would be possible by using point-by-point measurement methods.

Originally, the dc step-response technique was used to measure complex permittivity at very low frequencies, down to 10^{-4} Hz. Later, it was developed for use in the range 10^8–10^{10} Hz, the material being placed in a waveguide and subjected to a train of short fast-rise-time pulses. By observing the transmitted and reflected pulses with a sampling oscilloscope, the complex permittivity may be deduced. This technique is known as time-domain spectroscopy. Because of the way in which the reflections from the various interfaces are separated in time, a gating method can be used to isolate the required reflections, that is, those from the air-dielectric interfaces.

Automated instrumentation. The bridges described above are usually operated manually, but several automated systems have become available where the appropriate electrical properties of the cell

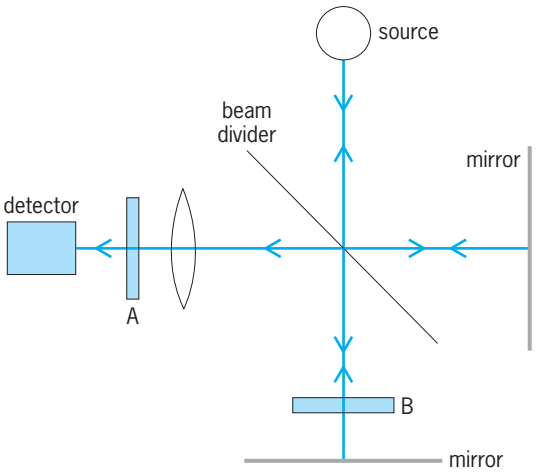

Fig. 8. Interferometer arrangement with specimen at A for conventional Fourier transform spectroscopy and at B for dispersive Fourier transform spectroscopy.

containing the dielectric material are measured automatically.

Automatic network analyzers (ANAs), developed originally for measuring the electrical properties of radio-frequency and microwave components, also measure a cell containing dielectric material, and the dielectric properties can be deduced from the difference in the electrical properties of the cell when empty and when containing the dielectric material.

Impedance analyzers can be used at low and radio frequencies. In general they are more accurate than ANAs for frequencies up to 200 MHz. In some cases the instruments are supplied with an external dielectric measurement cell and are often called materials analyzers.

Frequency response analyzers are instruments which are designed for dielectric measurements below 10 MHz down to as low as a few microhertz. The material is placed in an admittance cell which is in the form of a parallel-plate capacitor. The admittance is calculated from measurements of the in-phase and quadrature components of the voltage using an automated comparator circuit. The real and imaginary components of the permittivity of the material are obtained from the dimensions of the sample and the measured capacitance and conductance. These instruments have an impedance range from 0.01 ohm to 100 teraohms and a loss-angle resolution of less than 0.1 mrad. *See* ADMITTANCE; CAPACITANCE MEASUREMENT.

James H. Calderwood; R. Gareth Jones

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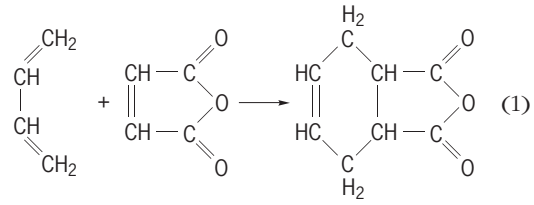
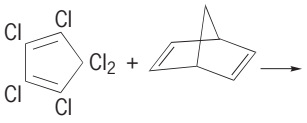
Diels-Alder reaction

The 1,4-addition of an alkene (the dienophile) to a conjugated diene. The reaction, also known as the diene synthesis, is one of the most valuable and versatile methods for the preparation of compounds containing a six-membered ring, and proceeds most rapidly when the dienophile is substituted by electron-attracting groups, such as —C═O or —C≡N [reactions (1) and (2)].

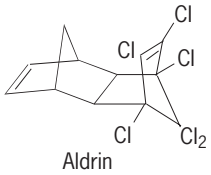


(1)

Diene(1,3-butadiene) Dienophile (maleic anhydride) Adduct



Hexachloro-cyclopentadiene Norborna-diene

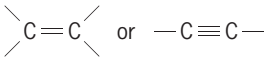


(2)

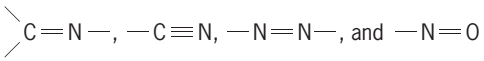
Aldrin

Diene component. With the exception of a few highly substituted diolefins which interfere with the reaction because of steric effects, most alkyl and aryl homologs of butadiene react readily. Alicyclic dienes are especially reactive, except in those cases where the adduct formed would contain a bridged ring having a double bond at the bridgehead. Aromatic compounds such as styrene, in which part of the conjugation is in the ring, react in a normal manner, but heterocyclic dienes may show anomalous behavior. Furan reacts normally to give a six-membered ring containing a bridged oxygen, whereas thiophene and pyrrole do not react in this manner.

Dienophile component. Most commonly, dienophiles consist of compounds containing structures



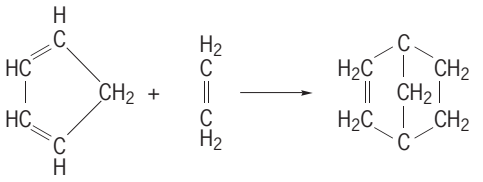
However, dienophiles are not restricted to unsaturated carbon compounds, and adducts form with



as the dienophile component. Quinones are especially reactive with dienes, requiring only mild conditions.

Reaction conditions. The Diels-Alder or diene-synthesis reaction does not require a catalyst, nor is the reaction retarded by the presence of oxidation inhibitors with which dienes are commonly treated to prevent formation of peroxides. For reactive dienophiles such as maleic anhydride, it is sufficient to mix the reactants in molar proportion, usually in a solvent such as benzene, and reaction takes

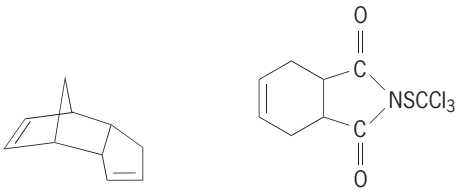
place at room temperature—often with the evolution of heat. On the other hand, the reaction of ethylene with cyclopentadiene to form bicyclo-[2.2.1]-2-heptene [reaction (3)] requires temperatures in the



(3)

range of 190–220°C (370–430°F) and pressures in the range of 20–80 atm (2–8 megapascals).

Instances are known where the diene synthesis is a reversible reaction. For example, cyclohexene, the adduct of 1,3-butadiene and ethylene, is quantitatively transformed into its components when subjected to high temperature for a short contact time at low pressure. Similarly, dicyclopentadiene (**1**) yields



(**1**) (**2**)

two moles of cyclopentadiene when maintained at its boiling point of 170°C (338°F).

Industrially the diene synthesis is used in the production of the insecticides aldrin and dieldrin. The adduct of butadiene and maleic anhydride is used in the synthesis of the fungicide captan (**2**).

Paul E. Fanta

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Diesel cycle

An internal combustion engine cycle in which the heat of compression ignites the fuel. Compression-ignition engines, or diesel engines, are thermodynamically similar to spark-ignition engines. The sequence of processes for both types is intake, compression, addition of heat, expansion, and exhaust. Ignition and power control in the compression-ignition engine are, however, very different from those in the spark-ignition engine.

Usually, a full unthrottled charge of air is drawn in during the intake stroke of a diesel engine. A compression ratio between 12 and 20 is used, in contrast to a ratio of 4 to 10 for the Otto spark-ignition engine. This high compression ratio of the diesel raises the temperature of the air during the compression stroke. Just before top center on the compression stroke, fuel is sprayed into the combustion chamber. The high temperature of the air ignites the

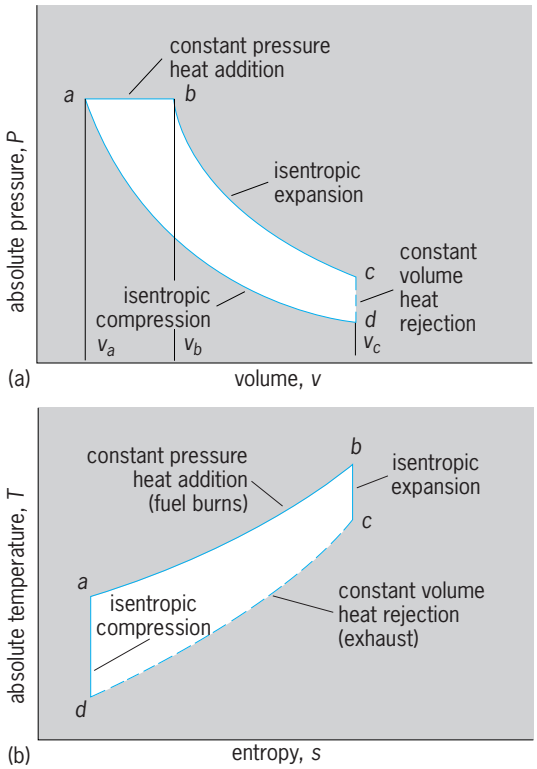

Fig. 1. Ideal diesel cycle, with (*a*) pressure-volume and (*b*) temperature-entropy bases.

fuel, which burns almost as soon as it is introduced, adding heat. The combustion products expand to produce power, and exhaust to complete the cycle.

Performance of a diesel engine is anticipated by analyzing the action of an air-standard diesel cycle. An insulated cylinder equipped with a frictionless piston contains a unit air mass. The metal cylinder head is alternately insulated and then uncovered for heat transfer.

Air is compressed until the piston reaches the top of the stroke. Then the air receives heat through the cylinder head and expands at constant pressure along path *a-b* as in **Fig. 1**, moving the piston part way down through the cylinder. Then the cylinder head is insulated, and the air completes its expansion along path *b-c* at constant entropy. The cylinder head is uncovered, and with the piston at the bottom of its stroke, a constant-volume heat rejection takes place on path *c-d*. The insulation is replaced, and the cycle is completed with an isentropic compression on path *d-a*.

An increase in compression ratio $r = v_d/v_a$ increases efficiency η, the increase becoming less at higher compression ratios. Another characteristic of the diesel cycle is the ratio of volumes at the end and at the start of the constant-pressure heat-addition process. This cutoff ratio $r_c = v_b/v_a$ measures the interval during which fuel is injected. For an engine to develop greater power output, the cutoff ratio is increased and heat continues to be added further into the expansion stroke. The air-standard cycle shows that, with less travel remaining during which to expend the additional heat energy as mechanical

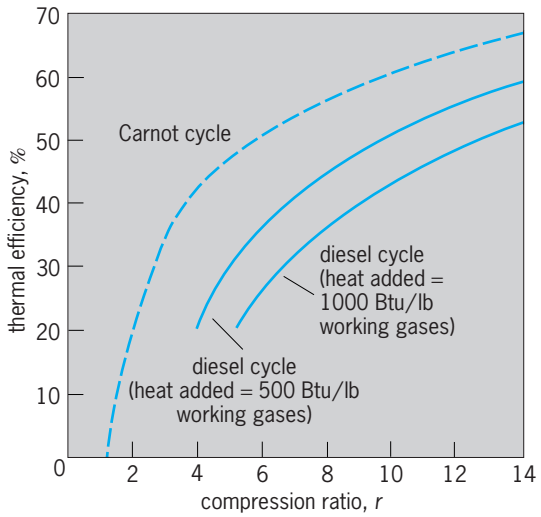

Fig. 2. Thermal efficiency of an ideal diesel cycle.

energy, the efficiency of the engine is reduced. Conversely, efficiency increases as the cutoff ratio decreases, so that a diesel engine is most efficient at light loads. Specifically, theequation below may be

$$\eta = 1 - \frac{1}{r^{k-1}}\left[\frac{r_c^k - 1}{k(r_c - 1)}\right]$$

written, where $k = c_p/c_v$, the ratio of specific heat of the working substance at constant pressure to its specific heat at constant volume (**Fig. 2**). In the limiting case when cutoff ratio r_c approaches unity, diesel cycle efficiency approaches Otto cycle efficiency for cycles of the same compression ratio.

In an actual engine with a given compression ratio, the Otto engine has the higher efficiency. However, fuel requirements limit the Otto engine to a compression ratio of about 10, whereas a diesel engine can operate at a compression ratio of about 15 and consequently at a higher efficiency.

In addition, heat can be added earlier in the cycle by injecting fuel during the latter part of the compression process *d-a*. This mode of operation is the dual-combustion or semidiesel cycle. With most of the heat added near peak compression, semidiesel efficiency approaches Otto cycle efficiency at a given compression ratio. *See* DIESEL ENGINE; INTERNAL COMBUSTION ENGINE.

Theodore Baumeister

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Diesel engine

An internal combustion engine operating on a thermodynamic cycle in which the ratio of compression ($R_v = 15\pm$) of the air charge is sufficiently high to ignite the fuel subsequently injected into the combustion chamber. Compared to an engine operating on the Otto cycle, the diesel engine utilizes a wider variety of fuels with a higher thermal efficiency and consequent economic advantage under many service applications (**Fig. 1**). *See* DIESEL CYCLE; OTTO CYCLE.

Features. The true diesel engine, as represented in most low-speed engines, uses a fuel-injection system where the injection rate is delayed and controlled to maintain constant pressure during combustion. Adaptation of this injection principle to higher engine speeds has necessitated departure from the constant-pressure specification, because the time available for fuel injection is so short (often 2 ms or less). Nonvolatile (distillate) fuels are burned to advantage in these engines, which cannot be rigorously identified as true diesels but properly should be called commercial diesels. However, all such engines are ordinarily classified as diesels. Performance data for selected diesel engines are given in **Table 1**.

Diesel engines give high intrinsic and actual thermal efficiency (20–40%); a sample comparative heat balance is shown in **Table 2**. The engine output is controlled by regulation of the fuel supplied but without variation of the air supply (100±% excess air at full load). Supercharging (10–15 lb/in.^{2} or 70–100 kilopascals) increases power output for a given cylinder size and engine speed. With two-cycle constructions, scavenging air (approximately 5 lb/in.^{2} or 35 kPa) is delivered by crankcase compression, front-end compression, or a separate rotary, reciprocating, or centrifugal blower. The engine cylinder may be without valves and with complete control of admission of scavenging air and release of spent gases in a two-port construction, the piston covering and uncovering the ports; or the cylinder may have a single port (for admission or release) uncovered by

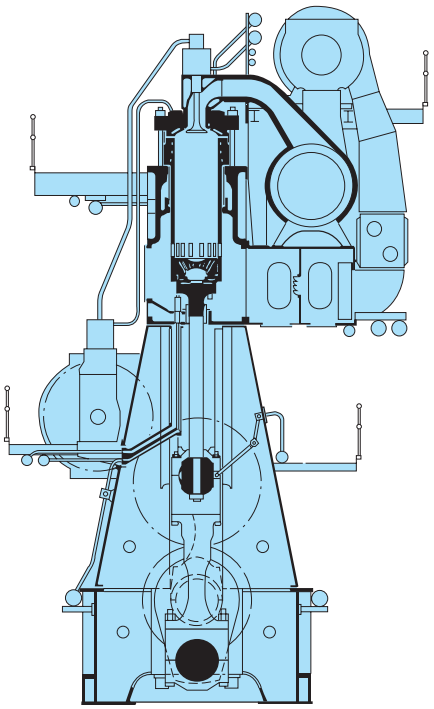

Fig. 1. Cross section of a slow-speed two-stroke diesel engine. (*Sulzer Brothers, Inc.*)

TABLE 1. Performance of selected diesel engines*

Type	Shaft horsepower (shp)	Ratio of compression, R_v	Brake mean effective pressure, lb/in.^{2}	Piston speed, ft/min	Mass, lb/hp	Overall thermal efficiency, %
Air injection engine	300–5000	12–15	50–75	600–1000	25–200	30–35
Solid injection, compression ignition						
Automotive	20–300	12–21	75–100	800–1800	7–25	25–30
Railroad	200–2500	12–15	60–90	800–1800	10–40	30–35
Stationary						
Unsupercharged	50–2500	12–15	70–80	600–1500	10–100	30–35
Supercharged	60–4000	10–13	110–125	600–1500	7.5–75	32–40
Dual fuel, stationary						
Unsupercharged	50–2500	12–15	80–90	600–1500	10–100	30–35
Supercharged	60–4000	10–13	120–135	600–1500	7.5–75	32–40

*1 hp = 746 W; 1 lb/in.^{2} = 6.9 kPa; 1 ft/min = 0.305 m/min ; 1 lb/hp = 6.08×10^{-4} kg/W.

the main piston at the outer end of its stroke and a cam-operated poppet valve in the cylinder head. The objective is to replace spent gases with fresh air by guided flow and high turbulence.

The four-cycle engine, with the intake and exhaust valves in the cylinder head, is most effective in scavenging. But the sacrifice of one power stroke out of every two is a frequent deterrent to its selection for use. Cylinder heads become complicated structures because of valve porting, jacketing, and injection-nozzle locations and the accommodation of these to effective combustion, heat transfer, and internal bursting pressures. The valve-in-head construction allows the camshaft to be located in the cylinder head or cylinder block.

Small-size (less than 200 horsepower or 150 kW) engines are conveniently started by an electric motor and storage battery. Larger engines use compressed air (about 200 lb/in.^{2} or 1380 kPa) introduced through valves in the cylinder head. Starting engine-driven generator sets may be accomplished by using the generator as a motor to crank the engine.

Cooling systems use water at 120–180°F (49–82°C) with radiators, cooling towers, and cooling ponds employed for conservation and reclamation. An antifreeze, such as ethylene glycol, and other additives are usually added to the water to prevent freezing and corrosion. Engine mountings and foundations must be designed to handle stress loadings and to reduce vibration. Exhaust systems should include wave-trap silencers or mufflers. Filters for air, fuel, lubricating oil, and (on some engines) cooling-system water help ensure engine reliability. *See* ANTIFREEZE MIXTURE; INTERNAL COMBUSTION ENGINE.

TABLE 2. Approximate allocation of losses in internal combustion engines

Type of loss	Otto engines, %	Diesel engines, %
Output	20	33
Exhaust losses	40	33
Cooling system losses	40	33
Other	<1	1
Total (input)	100	100

Passenger-car engines. Diesel engines have been installed in production automobiles in Europe since 1936. However, their acceptance in the United States was limited until after the oil shortages of 1973 and 1979, when some American manufacturers introduced diesel-powered automobiles to provide an option for improved fuel economy—generally about 25% better than that of a gasoline engine of similar displacement. Easing of the oil shortage in the early 1980s along with an increase in the price of diesel fuel contributed to a decline in public acceptance of the diesel engine.

The diesel engine in the automobile is usually a four-stroke-cycle engine with indirect injection into an auxiliary combustion chamber (**Fig. 2**). Most automobile diesel engines use a distributor-type injection pump. The fuel system often includes a fuel-conditioner assembly, which combines a water-in-fuel detector, water-fuel separator, fuel filter, fuel heater, and hand-priming pump in a single unit. *See* COMBUSTION CHAMBER.

The passenger-car diesel engine may be naturally aspirated, turbocharged, or turbocharged and intercooled. Some automobile diesel engines operate at speeds of up to 5000 rpm. Although the compression ratio with indirect injection ranges from about 20:1 to 23:1, cold starting requires additional heat from an electric glow plug located in each auxiliary chamber. *See* TURBOCHARGER.

The automotive diesel engine usually has four to six cylinders arranged in-line, or six or eight cylinders arranged in a V. Engine displacement ranges from about 1.5 to 7 liters. Some automotive four-cycle gasoline engines have been converted to diesel operation but, in general, these engines have not been as successful in the marketplace as an engine designed as a diesel.

Truck and bus engines. Diesel engines in trucks and buses are usually larger and operate at lower speeds than diesel engines in passenger cars. The typical truck and bus engine ranges in size from about 5 to 19 liters displacement, has a maximum speed of 3000 rpm or less, has direct injection into an open combustion chamber, and may be naturally aspirated, turbocharged, or turbocharged and aftercooled.

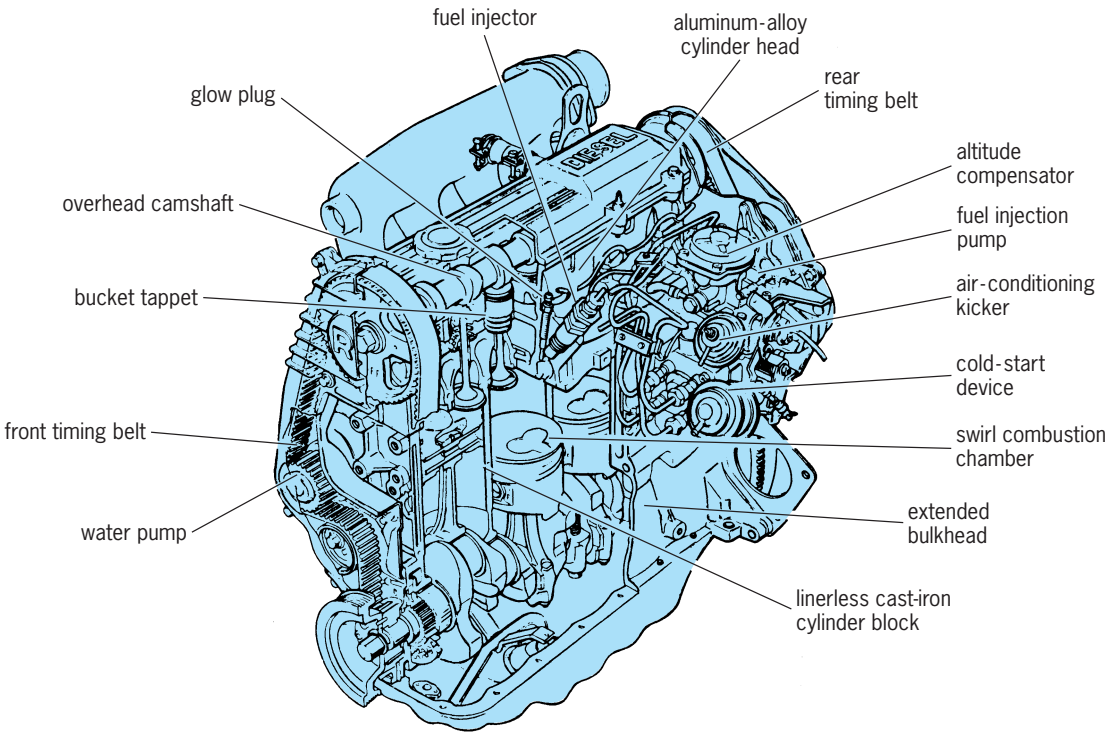

Fig. 2. A 2.0-liter, four-cylinder, four-stroke-cycle passenger-car diesel engine which has a distributor-type injection pump and indirect injection. (*Ford Motor Co.*)

Most truck diesel engines operate on the four-stroke cycle, although many buses and some trucks have two-stroke-cycle engines. These usually have intake ports in the cylinder and exhaust valves in the cylinder head, with scavenging air provided by a crankshaft-driven blower mounted on the crankcase. A unit fuel injector operated by the engine camshaft meters and injects the fuel into the combustion chamber at high pressure at the proper time.

Emissions. In addition to a noticeable odor, the exhaust gas from the diesel engine contains gaseous and particulate emissions which contribute to air pollution. The particles, or soot, may be removed by a trap oxidizer that consists of a filter and a regeneration system, which burns the trapped particles and cleans the filter. Gaseous emissions of unburned hydrocarbons (HC), carbon monoxide (CO), and oxides of nitrogen (NO_x) may be controlled through the use of electronically controlled fuel injection, exhaust-gas recirculation, and charge-air cooling. Operating the engine on low-sulfur fuel reduces sulfur and particulate emissions. *See* AIR POLLUTION; DIESEL FUEL; FUEL INJECTION.

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Diesel fuel

A broad mixture of hydrocarbons produced as distillates, as residual materials, or as blends of the two during the refining of crude petroleum. Diesel fuel usually has a distillation range of 390–715°F (200–380°C) and a specific gravity range of 0.760–0.935 [equivalent to 61.2–19.8° on the American Petroleum Institute (API) scale]. In addition to these properties, diesel fuel must have <1 weight % sulfur, < 0.1 weight % ash, < 0.5 volume % water and sediment, and a high flash point (greater than 131°F or 55°C).

Cetane number. Diesel fuel quality is defined by the cetane number, which usually falls into the range 30–60. A high cetane number indicates the potential for easy starting and smooth operation of the engine. The cetane number is the analog of the automobile engine octane number, with cetane (*n*-hexadecane, $C_{16}H_{34}$) having the arbitrarily assigned number of 100. At the other end of the scale, heptamethylnonane, an isomer of cetane, has the assigned cetane number of 0. The cetane number of a diesel fuel is determined by comparison with blends of cetane and heptamethylnonane. It corresponds to the number of parts by volume of cetane in a cetane-heptamethylnonane blend which has the same ignition quality as the fuel. *See* CETANE NUMBER.

For general comparison, the cetane number of paraffinic hydrocarbons, particularly those with straight rather than branched chains, is high, whereas the cetane numbers of aromatic and hydroaromatic (naphthenic) hydrocarbons is low. A fuel having a high octane number tends to have a low cetane number, and vice-versa. Diesel fuels usually contain 60–80 volume % paraffinic materials and 20–40 volume % aromatic-naphthenic materials. Diesel fuels having low cetane numbers (that is, around 60 volume % paraffinics) can be improved by the addition of chemicals such as amyl nitrate ($C_5H_{11}NO_3$) or hexyl nitrate ($C_6H_{13}NO_3$).

Categories. The American Society for Testing and Materials (ASTM) has categorized diesel fuels into three general groups. The need to categorize these fuels results from the varied uses of diesel engines, which are designed to operate efficiently on one of the standard diesel fuels. Use of a fuel which does not meet the designated specifications could have a detrimental effect on engine efficiency and life, and could cause increased emissions and noise (low cetane number), engine wear, and corrosion (high sulfur, high ash, and high sediment). *See* DIESEL ENGINE.

No. 1-D. This is a light distillate, similar to kerosine, for engines where frequent load changes and speed changes (truck, tractor engines) are essential. This fuel has a flash point greater than 100°F (38°C), with a minimum cetane number of 40. This fuel is believed to be particularly suitable for cold-weather operation. *See* KEROSINE.

No. 2-D. A medium distillate fuel with a lower volatility and higher density than No. 1-D, this fuel finds use in heavier-duty engines, for example, railroad engines, which operate at uniform speeds but with heavier loads than encountered during the use of No. 1-D. The flash point is greater than 125°F (52°C) and the minimum cetane number is 40.

No. 3-D. This is a heavy distillate fuel with the highest density and lowest volatility of the three diesel fuels. It finds use in low- and medium-speed engines such as marine engines and electric power generation engines, which operate under sustained loads. The flash point is greater than 130°F (54°C) and the minimum cetane rating is 30.

Other specifications. The result of so-called energy crises was the use of diesel engines in automobiles and the increased use of diesel engines in public transportation vehicles such as buses. These engines require diesel fuel with the highest possible cetane number (60 or more) and could even result in the identification and classification of an additional grade of diesel fuel. *See* PETROLEUM PRODUCTS.

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Difference equation

A relationship between one or more independent variables, one or more dependent variables, and differences of those variables. Difference equations arise in the analysis of discrete systems (for example, a string loaded along its length with small masses), in the solution of differential equations by means of digital computers, in the implementation of digital filters, and in the discrete-time control of systems. An ordinary difference equation expresses a relationship between an independent variable t and one or more dependent variables, $y(t)$, $w(t)$, and so forth, and any successive differences of y, w, ... The first forward difference of y, relative to the increment h, is defined by Eq. (1). The second forward difference

$$\Delta y(t) = y(t+h) - y(t) \tag{1}$$

can be obtained by applying the above definition to itself, as in Eq. (2).

$$\begin{aligned}\Delta^2 y(t) &= \Delta[\Delta y(t)] \\ &= \Delta y(t+h) - \Delta y(t) \\ &= [y(t+2h) - y(t+h)] \\ &\quad -[y(t+h) - y(t)] \\ &= y(t+2h) - 2y(t+h) + y(t)\end{aligned} \tag{2}$$

A difference equation is linear if it is of the first degree with respect all the quantities $\Delta y(t)$, $\Delta^2 y(t)$, $\Delta^3 y(t)$, ... The general solution of difference equations is possible only for the case where they are linear with constant coefficients. For this reason and because of their importance, attention will be initially limited to linear, constant-coefficient difference equations.

Linear constant-coefficient equations. The independent variable will usually have discrete values. Then the notation can be simplified by making the substitution in Eq. (3). In this case, the arguments of

$$\begin{aligned}t_k &= t_0 + kh \\ k &= 0, 1, 2, \ldots\end{aligned} \tag{3}$$

the independent variables, if more than one, can be replaced by the subscript k. Then, the forward difference operators given by Eqs. (1) and (2) can be written as Eqs. (4), where the shifting operator E has

$$\begin{aligned}\Delta y_k &= y_{k+1} - y_k = Ey_k - y_k \\ \Delta^2 y_k &= E^2 y_k - 2Ey_k + y_k\end{aligned} \tag{4}$$

been defined, incrementing the index on the dependent variable by 1 each time it is applied.

An illustration is given by the first-order difference equation (5) with initial condition $y_0 = 1$. It can be

$$y_{k+1} - ay_k = 1 \text{ or } Ey_k - ay_k = 1 \tag{5}$$

$$k \geq 0, a \text{ constant}$$

solved iteratively by putting the term ay_k on the right-hand side and finding the value of y_{k+1} in terms of y_k. Proceeding in this fashion, Eqs. (6) are obtained, or, in general, Eq. (7).

$$\begin{aligned}y_0 &= 1 \\ y_1 &= 1 + a \\ y_2 &= 1 + a + a^2 \\ &\ldots\end{aligned} \tag{6}$$

$$y_k = 1 + a + a^2 + \cdots + a^k = \frac{1 - a^{k+1}}{1 - a} \tag{7}$$

$$a \neq 1$$

The general homogeneous difference equation of order n can be written in terms of the shifting

operator as Eq. (8). From the solution of linear,

$$E^n y_k + A_1 E^{n-1} y_k + \cdots + A_{n-1} E y_k + A_n y_k = 0 \quad (8)$$
$$k \geq 0$$

constant-coefficient differential equations, the solution is assumed to be of the form $e^{rk} = \beta^k$, where r is a suitably chosen constant and $\beta = e^r$. From the definition of the shifting operator, Eqs. (9) follow, as well

$$\begin{aligned} E\beta^k &= \beta^{k+1} = \beta(\beta^k) \\ E^2\beta^k &= \beta^{k+2} = \beta^2(\beta^k) \end{aligned} \quad (9)$$

as similar equations for heigher powers of E. When these equations are substituted into Eq. (8), the result is Eq. (10). Hence, $y_k = \beta^k$ will satisfy Eq. (8) if β is a

$$\beta^n + A_1\beta^{n-1} + \cdots + A_{n-1}\beta + A_n = 0 \quad (10)$$

root of Eq. (10). If all roots are real and distinct, the most general expression satisfying Eq. (8) is Eq. (11),

$$y_k = c_1\beta_1^k + c_2\beta_2^k + \cdots + c_n\beta_n^k \quad (11)$$
$$k \geq 0$$

where $\beta_1, \beta_2, \ldots, \beta_n$ are the roots to Eq. (10) and c_1, $c_2, \ldots, c_n$ are constants to be determined by the initial conditions. *See* ALGEBRA.

An example is given by the second-order linear difference equation (12). In this case Eq. (10) becomes

$$E^2 y_k - 3E y_k + 2y_k = 0 \quad (12)$$

Eq. (13), with solutions $\beta = 2$ and $\beta = 1$. Thus, the

$$\beta^2 - 3\beta + 2 = 0 \text{ or } (\beta - 2)(\beta - 1) = 0 \quad (13)$$

solution to Eq. (12) is Eq. (14), which follows since $1^k = 1$.

$$y_k = c_1 2^k + c_2 \quad (14)$$
$$k \geq 0$$

If a root is repeated, for example, β_1, then the two solutions corresponding to this repeated root are of the form β_1^k and $k\beta_1^{k-1}$. In the case of the difference equation (15), the corresponding roots of Eq. (10)

$$E^3 y_k - 3E^2 y_k + 4y_k = 0 \quad (15)$$

are -1, 2, and 2, and the solution is then Eq. (16).

$$y_k = c_1(-1)^k + (c_2 + c_3 k)2^k \quad (16)$$
$$k \geq 0$$

Complex roots for Eq. (10) will always occur in complex conjugate pairs for a difference equation with real coefficients. When this occurs, these complex conjugate roots, for example, β_1 and β_2, can be written in polar form as $\rho e^{j\phi}$ and $\rho e^{-j\phi}$ (where $j = \sqrt{-1}$), and the corresponding solution is equivalent to Eq. (17).

$$y_k = \rho^k(c_1 \cos k\phi + c_2 \sin k\phi) \quad (17)$$
$$k \geq 0$$

See COMPLEX NUMBERS AND COMPLEX VARIABLES.

Nonhomogeneous linear equations. The general solution of a nonhomogeneous difference equation, given by Eq. (18), is of the form given by Eq. (19),

$$E^n y_k + A_1 E^{n-1} y_k + A_{n-1} E y_k + A_n y_k = x_k \quad (18)$$
$$k \geq 0$$

$$y_k = y_{k,h} + y_{k,p} \quad (19)$$

where $y_{k,h}$ is the solution of the homogeneous equation [$x_k = 0$ in Eq. (18)], and $y_{k,p}$ is the particular solution of Eq. (18). With constant coefficients on Eq. (18), the method of undetermined coefficients can be used to find the particular solution. In this procedure, three steps are followed in constructing the form of the particular solution:

1. The family of a term represented on the right-hand side of the difference equation (called the forcing function) is constructed. (The family of a term f_k is defined to be the set of all functions of which f_k and all operations $E^m f_k$ are linear combinations. For example, for the forcing function $x_k = e^{bk}$, b a constant, these operations are $E^m e^{bk} = (e^{bm})e^{bk} =$ constant $\times\ e^{bk}$.
2. If that family has no representatives in the homogeneous solution, $y_{k,p}$ is assumed to be a linear combination of the members of that family, and constants of the combination are determined so that the difference equation is identically satisfied.
3. If that family has a representative is the homogeneous solution, each member of the family is multiplied by the smallest integer power of k for which all such representatives are removed, and the particular solution is assumed to be a linear combination of the members of the modified family.

For example, the difference equation (20) corre-

$$E^2 y_k - 3E y_k + 2y_k = 1 + \alpha^k \quad (20)$$
$$k \geq 0$$

sponds to the homogeneous equation (12) with the homogeneous solution given by Eq. (14). The particular solution is assumed to have the form given by Eq. (21), since 1 is a solution of the homogeneous

$$y_{k,p} = Ak + B\alpha^k \quad (21)$$

equation (and is therefore multiplied by k). Substitution into Eq. (20) gives conditions for the undetermined coefficients A and B which result in Eqs. (22).

$$\begin{aligned} A &= -1 \\ B &= \frac{\alpha}{(\alpha - 1)(\alpha - 2)} \end{aligned} \quad (22)$$

When these equations are substituted into Eq. (21) and the particular solution is used in Eq. (19) along with the homogeneous solution found earlier, the total solution is found to be given by Eq. (23).

$$Y_k = c_1 2^k + c_2 - k + \frac{\alpha^{k+1}}{(\alpha - 1)(\alpha - 2)} \quad (23)$$
$$k \geq 0$$

The solution of constant-coefficient, linear difference equations is facilitated greatly by use of the z transform. *See* TRANSFORM.

Nonlinear difference equations. Nonlinear difference equations arise naturally in the solution of nonlinear differential equations. For simplicity, a first-order ordinary differential equation of the form given by Eq. (24) will be considered. If it appears impossible to integrate Eq. (24), a numerical solution may be attempted by starting with the value y_0 at t_0 and computing the approximate solution y_1 at time $t_0 + h$, the approximate value y_2 at time $t_2 = t_0 + 2h$ and so on. By a Taylor series expansion, Eq. (25), holds,

$$\frac{dy(t)}{dt} = f(t, y) \tag{24}$$

$$y(t_0) = y_0$$

$$y(t+h) = y(t) + hy'(t) + \frac{h^2}{2}y''(t) + \cdots$$

$$= y(t) + hf(t, y) + \frac{h^2}{2}f'(t, y) + \cdots \tag{25}$$

where the prime indicates differentiation with respect to t and Eq. (24) has been used to get the second equation. For small values of h, a crude approximation is given by the first two terms of Eq. (25), as in Eq. (26). By using the index notation introduced

$$y(t+h) = y(t) + hf \tag{26}$$

earlier, the solution to Eq. (24) can be approximated by the numerical algorithm given by Eq. (27). This

$$y_{k+1} = y_k + hf(k, y_k) \tag{27}$$

$$k = 0, 1, 2, \ldots$$

is called the Euler method for solving the differential Eq. (24). Because the expansion of Eq. (25) for $y(t + h)$ is truncated at the term linear in h, the Euler method referred to as first order. For small h, the error induced by this truncation is of the order of h^2. *See* DIFFERENTIATION.

An improvement in the accuracy of the Euler method is provided by the modified Euler method, which is expressed by Eqs. (28). This is called the improved Euler method, and can be viewed as approximating the integration by a straight line through the point (t_k, y_k) with slope $f(t_k, y_k)$ in the first half interval and by a straight line of slope $f(t_{k+1}, y^*_{k+1})$ in the second half interval. Thus, it is called a predictor-corrector method, and has a truncation error of the order of h^3.

$$\begin{aligned} y^*_{k+1} &= y_k + hf(t_k, y_k) \\ y_{k+1} &= y_k + {}^1\!/_2 h[f(t_k, y_k) + f(t_{k+1}, y^*_{k+1})] \end{aligned} \tag{28}$$

The Runge-Kutta method is a fourth-order method and has truncation error of the order h^5. It is implemented by Eqs. (29).

$$\begin{aligned} k_1 &= hf(t_k, y_k) \\ k_2 &= hf(t_k + {}^1\!/_2 h, y_k + {}^1\!/_2 k_1) \\ k_3 &= hf(t_k + {}^1\!/_2 h, y_k + {}^1\!/_2 k_2) \\ k_4 &= hf(t_k + h, y_k + k_3) \\ y_{k+1} &= y_k + {}^1\!/_6(k_1 + 2k_2 + 2k_3 + k_4) \end{aligned} \tag{29}$$

The above integration methods are one-step in that the integration is based on values of $f(t, y)$ in a single interval. Multistep methods provide even greater accuracy but require values for y_k from more than one interval. Therefore, one of the challenges in starting them is to provide the necessary initial values, which is typically done with a one-step method.

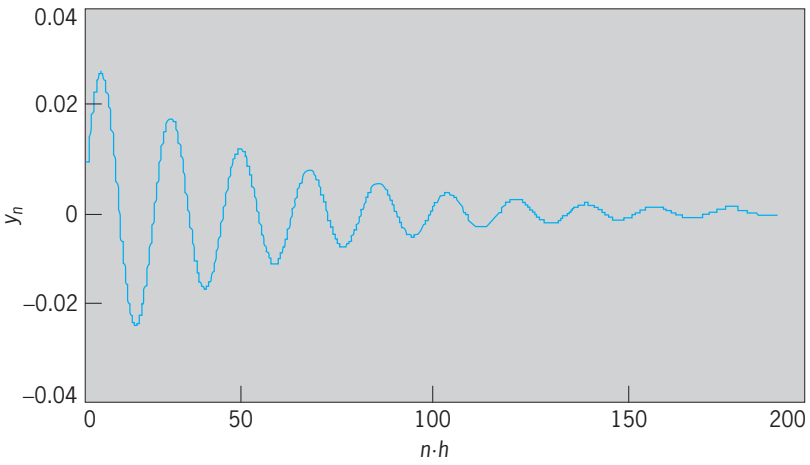
Methods also exist for integrating nonlinear differential equations of order higher than one. For example, for the second-order differential equation (30), an algorithm may be based on Eqs. (31), where the prime denotes differentiation with respect to time and the value for y_k can be substituted from Eq. (30). A more accurate method is provided by a generalization of the Runge-Kutta method.

$$y''(t) = f(t, y, y') \tag{30}$$

$$\begin{aligned} y_{k+1} &= y_k + hy'_k + \frac{h^2}{2}y''_k \\ y'_{k+1} &= y'_k + hy''_k \end{aligned} \tag{31}$$

Solutions of equations for nonlinear systems based on the above integration algorithms are ill-posed in that an incorrect steady-state value may be reached. Alternative schemes based on state variables avoid this problem and are faster.

Approximating continuous-time systems. Methods exist for approximating governing equations for linear continuous-time systems by difference equations. For example, a continuous-time second-order linear system represented by the differential



Numerical solution of Eq. (32) for $a = 0.05$, $b = 0.1$, $c = 0.01$, and a step input. The solution uses the state-variable representation of Eq. (33) and the difference-equation approximation of Eq. (36) with $h = 0.1$.

equation (32) can be recast in state-variable form as Eqs. (33). The state equations [the first two equations

$$\frac{d^2y(t)}{dt^2} + a\frac{dy(t)}{dt} + by(t) = c\frac{dx(t)}{dt} \tag{32}$$

$$\begin{aligned} \frac{dq_1(t)}{dt} &= -aq_1(t) - q_2(t) + cx(t) \\ \frac{dq_2(t)}{dt} &= bq_1(t) \\ y(t) &= q_1(t) \end{aligned} \tag{33}$$

of Eqs. (33)] can be written in matrix form as Eq. (34), where the matrices **A** and **B** are given by Eqs. (35).

$$\dot{\mathbf{Q}}(t) = \mathbf{AQ}(t) + \mathbf{Bx}(t) \tag{34}$$

$$\mathbf{A} = \begin{bmatrix} -a & -1 \\ b & 0 \end{bmatrix} \qquad \mathbf{B} = \begin{bmatrix} c \\ 0 \end{bmatrix} \tag{35}$$

Approximation of these equations using Euler integration results in Eqs. (36), which can be used to obtain a numerical solution (see **illus.**).

$$\begin{aligned} q_{1,k+1} &= (1 - ha)q_{1,k} - hq_{2,k} + hcx_k \\ q_{2,k+1} &= jhbq_{1,k} + q_{2,k} \\ y_k &= q_{1,k} \end{aligned} \tag{36}$$

Partial difference equations. The numerical solution of a partial differential equation leads to a partial difference equation. The procedure is similar to that used for numerical solution of ordinary differential equations. The Laplace equation (37) is an exam-

$$\nabla^2 u = \frac{\partial^2 u}{\partial x^2} + \frac{\partial^2 u}{\partial y^2} = 0 \tag{37}$$

ple of a simple partial differential equation. By using Taylor's expansion, the first partial derivatives of u with respect to x and y can be approximated by Eqs. (38). Similarly, the second derivatives can be

$$\begin{aligned} \frac{\partial u(x,y)}{\partial x} &\approx \frac{1}{2h}[u(x+h,y) - u(x-h,y)] \\ \frac{\partial u(x,y)}{\partial y} &\approx \frac{1}{2k}[u(x,y+k) - u(x,y-k)] \end{aligned} \tag{38}$$

approximated by Eqs. (39). Substitution into the

$$\begin{aligned} \frac{\partial^2 u(x,y)}{\partial x^2} &\approx \frac{1}{h^2}[u(x+h,y) - 2u(x,y) \\ &\qquad + u(x-h,y)] \\ \frac{\partial^2 u(x,y)}{\partial y^2} &\approx \frac{1}{k^2}[u(x,y+k) - 2u(x,y) \\ &\qquad + u(x,y-k)] \end{aligned} \tag{39}$$

Laplace equation yields the difference Eq. (40). In

$$\begin{aligned} u(x+h,y) + u(x,y+h) + u(x-h,y) \\ + u(x,y-h) - 4u(x,y) = 0 \end{aligned} \tag{40}$$

the case of the Poisson equation, which is Laplace's equation with the right-hand-side 0 replaced by the known function $f(x, y)$, the difference equation (40) has the right-hand side replaced by $h^2 f(x, y)$. In Eq. (40), h is called the mesh size, and the equation itself relates $u(x, y)$ to the values of u at the four nearest-neighbor positions. In fact, for the Laplace equation, u at (x, y) equals the mean of the values at the four neighboring positions. *See* DIFFERENTIAL EQUATION; DIGITAL FILTER; INTERPOLATION; LAPLACE'S DIFFERENTIAL EQUATION; LINEAR SYSTEM ANALYSIS; NUMERICAL ANALYSIS. Rodger E. Ziemer

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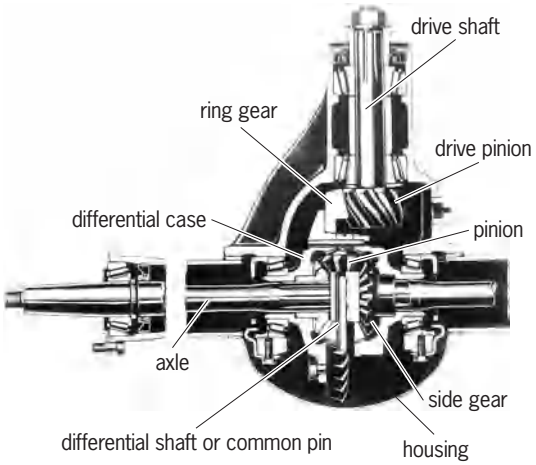
Differential

A mechanism which permits a rear axle to turn corners with one wheel rolling faster than the other. An automobile differential is located in the case carrying the rear-axle drive gear (see **illus.**).

The differential gears consist of the two side gears carrying the inner ends of the axle shafts, meshing with two pinions mounted on a common pin located in the differential case. The case carries a ring gear driven by a pinion at the end of the drive shaft. This arrangement permits the drive to be carried to both wheels, but at the same time as the outer wheel on a turn overruns the differential case, the inner wheel lags by a like amount.

Special differentials (Dana, Borg Warner, Eaton) permit one wheel to drive the car by a predetermined amount even though the opposite wheel is on slippery pavement; they have been used on racing cars for years and are now used by a number of car manufacturers.



Commonly used rear-axle differential. (*Chrysler Corp.*)

In operation, engine torque applied to the differential case causes the angular contacts on the case to bear on corresponding angles on the differential pinion pins. Two contacts 180° apart push one pin to the right; the other two contacts, 180° apart, spaced 90° from the first, push the other pin to the left. On one of the pins are two pinions meshing with the right-side gear; on the other are two pinions meshing with the left-side gear. Diameters on the pinions concentric with the teeth bear against the side gear rings splined to the inner end of the axle shafts and force the rings to apply clutches which connect the differential case directly to the rings. The pinions also load the side gears axially, the latter contributing their thrusts to the clutches. The drive thus passes from the differential case to the rings splined to the respective rear-axle shafts, providing a frictional drive to one wheel, even though the other is on ice or slippery pavement. This nullifies the conventional differential action by the predetermined amount. *See* AUTOMOTIVE TRANSMISSION; GEAR. Harold Fischer

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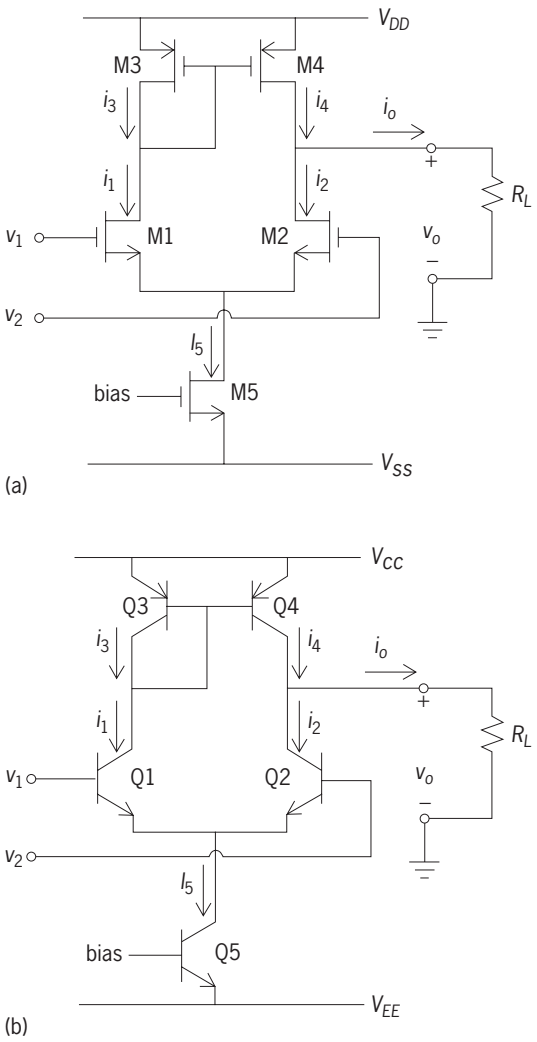
Differential amplifier

An electronic circuit that is designed to amplify the difference between two voltages measured with respect to a common reference, usually designated as ground. By convention, the net difference of two voltages measured with respect to a common reference is called the differential-mode voltage, while the sum of the voltages, usually divided by two to give an average value, is called the common-mode voltage.

An ideal differential amplifier thus has exactly the same gain from each input to its output, and the amplifier produces an output that is directly proportional to its differential-mode voltage. The amplifier delivers zero output in response to common-mode voltages. If these gains are not exactly equal, then equal (common-mode) voltages applied at each input terminal will not be equal at the amplifier output and their difference will not cancel completely. The common-mode gain, the ratio of the output response of a real differential amplifier to the input signal applied equally to each input terminal, is a measure of this gain mismatch.

Differential amplification is very useful when the signal to be amplified exists in an electrically noisy environment, since the noise voltage is usually a common component of both input voltages and, hence, will cancel when the difference of the amplifier inputs is taken.

Operation. For a physical differential amplifier to work properly, the electrical paths of each input signal through the amplifier must be nearly identical. Thus, the most important requirement for a differential amplifier is that it be constructed with transistors with closely matched electrical characteristics. Integrated circuits with amplifier transistors physically



Differential amplifier using (*a*) metal-oxide-semiconductor field-effect transistors (MOSFETs), M1–M5, and (*b*) bipolar junction transistors (BJTs), Q1–Q5. V_{DD}, V_{CC} = positive power supplies; V_{SS}, V_{EE} = negative power supplies; v_1, v_2 = input voltages; M1(Q1), M2(Q2) = input transistors with drain (collector) currents i_1, i_2; M3(Q3), M4(Q4) = load transistors forming current mirror with equal source (emitter) currents, i_3 and i_4; transistor M5(Q5) functions as current sink with value I_5; I_O = output current; v_O = output voltage; R_L = load resistance.

close to each other meet the required close matching requirement and are ideally suited for the production of differential amplifiers. *See* INTEGRATED CIRCUITS.

Differential-amplifier circuits that are suitable for integrated-circuit fabrication can use either metal-oxide-semiconductor field-effect transistors (MOSFETs; **illus.** *a*) or bipolar junction transistors (BJTs; illus. *b*). The input transistor pair must be matched closely. For best performance, the two load transistors also should be matched. *See* TRANSISTOR.

A fifth transistor functions as a dc current sink whose value, I_5, is determined by the voltage applied to the gate (base) of this transistor. The positive and negative power supplies are normally equal in magnitude. When the input voltages are equal, the drain (collector) currents of the input transistors are

identical and equal to exactly half of the dc current sink. Since the source (emitter) current of the left load transistor equals the drain (collector) current of the left input transistor, it follows that if the former current is half of the dc current sink, the latter current has this same value. The function of the load transistor pair is to provide two equal currents [the source (emitter) currents of these transistors]. This circuit is called a current mirror because it forces one of these currents to equal the other. Under the above conditions, both the drain (collector) current of the right input transistor and the source (emitter) current of the corresponding member of the current-mirror pair equal half the dc current sink. These two currents leave and enter the output terminal, respectively, and therefore no current flows through the output (and the output voltage is zero). *See* CURRENT SOURCES AND MIRRORS.

If the left input voltage is increased and the right remains constant, part of the increase in the left input voltage will cause an increase in the drain (collector) current of the left input transistor, to which this voltage is applied. The other part will cause a decrease in the drain (collector) current of the right input transistor. The mirroring function of the load transistor pair forces the source (emitter) current of the right load transistor to equal that of the left, which in turn equals the increased drain (collector) current of the left input transistor. Summing currents at the output terminal of the differential amplifier shows that the output current must equal the sum of the increase in the drain (collector) current of the left input transistor and the decrease in the corresponding current of the right input transistor. If the left input voltage is decreased or the right input voltage is increased, the changes become negative and the output current becomes equal to the negative of this sum.

Gain. The output current flows through the resistance seen at the output terminal, R_1, to create the output voltage of the differential amplifier. This output voltage is referenced to ground and is called the single-ended output voltage. The maximum output resistance is equivalent to the parallel combination of the small-signal drain-source resistances of the right input transistor or the collector-emitter resistances of the corresponding current-mirror transistor. Typical voltage gains from the differential input to the signal-ended output range from 50 to 200 for the MOSFET differential amplifier and 500 to 2000 for the BJT differential amplifier. The BJT differential amplifier has a larger gain because the transconductance of the BJT is approximately 10 times greater than the transconductance of the MOSFET. The gain of the BJT differential amplifier can be increased by using cascode transistors between the right input transistor and the corresponding current-mirror transistor and the output. *See* CASCODE AMPLIFIER; GAIN.

Frequency response. The frequency response of the differential amplifier is dependent on the reciprocal product of the output resistance and the capacitance in parallel with this output resistance. Typical values of output resistance range from 50 to 500 kΩ. For a 10-picofarad capacitor connected to the output, the frequency where the gain has decreased by −3dB from its low-frequency value ranges from 32 to 320 kHz. As the load capacitor increases, this frequency will decrease. *See* RESPONSE.

Limitations. The above considerations are appropriate when the current changes discussed above are small compared to the dc current sink. When large changes are applied to either or both of the input voltages, all the dc current sink will flow through one side of the differential amplifier only. For example, if the left input voltage is much larger than the right, the drain (collector) current of the left input transistor is equal to the dc current sink and the corresponding current of the right input transistor is equal to zero. Therefore, the current mirror forces the output current to equal the dc current sink. Similarly, if the left input voltage is much less than the right, the output current becomes the negative of the dc current sink. Thus, the maximum sinking or sourcing current of the differential amplifier output is $+I_5$. When a load capacitor, C_L, is attached to the output, the voltage change rate will be limited to a value of $+I_5/C_L$. This rate limitation is called slew rate and depends on the values of I_5 and C_L.

An important characteristic of the differential amplifier is the input common-mode voltage range. This is the common-mode voltage over which the differential amplifier can still function as a differential amplifier without any degradation. The upper limit is the common-mode voltage at which the current-mirror transistors no longer operate as a current mirror because the voltage across them becomes too small. The lower limit depends on the bias setting of the transistor that functions as a dc current sink for the same reason. A differential amplifier will have an input common-mode voltage range of about 2–3V less than the difference between the positive and negative power-supply voltages.

If the output changes when the same signal is applied to the two input voltages, the common-mode gain of the differential amplifier is not zero. The causes of common-mode gain are mismatches between the input transistors, imperfect current mirroring, and a finite value of resistance of the dc current sink implemented by the fifth transistor. If the output is not zero when a dc voltage is applied to both the input voltages, the differential amplifier has a dc offset. This dc offset at the output is divided by the differential to single-ended voltage gain and reflected to the input of the differential amplifier, where it is called the input offset voltage. The input offset voltage can be as large as 15 mV for MOSFET differential amplifiers and 5 mV for BJT differential amplifiers. *See* AMPLIFIER; OPERATIONAL AMPLIFIER.

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Differential equation

A relationship between a function and its derivatives.

If there is one independent variable, the differential equation is called an ordinary differential equation. The general form of such an equation is shown in Eq. (1), where t is the independent variable; u

$$F(t, u, u', u'', \ldots, u^{(n)}) = 0 \tag{1}$$

is a function of t; $u' = du/dt$, $u'' = d^2u/dt^2$, ..., $u^n = d^nu/dt^n$ are the derivatives of u; and $F(t, u_0, u_1, \ldots, u_n)$ is a given function of the $n + 2$ variables $t, u_0, \ldots, u_n$. The positive integer n is called the order of the differential equation; that is, the order of the differential equation is the order of the highest derivative that occurs in the equation. As an example, consider Eq. (2), which is an or-

$$u'' + u = 0 \tag{2}$$

dinary differential equation of order 2. A function u is said to be a solution to Eq. (1) if, when u and its derivatives up to order n are substituted into F, the identity expressed in Eq. (1) is valid for t in some interval $a < t < b$. In Eq. (2), $u = \cos t$ and $u = \sin t$ are solutions for $-\infty < t < \infty$, as can be immediately verified.

If there are two or more independent variables, the equation is called a partial differential equation. Again, the order of the equation is the order of the highest partial derivative that occurs in the equation. Thus, the general form of a partial differential equation of order 2, with two independent variables, is given by Eq. (3), where x and y are the indepen-

$$F(x, y, u, u_x, u_y, u_{xx}, u_{xy}, u_{yy}) = 0 \tag{3}$$

dent variables; $u = u(x,y)$ is a function: $u_x = \partial u/\partial x$, $u_y = \partial u/\partial y$, $u_{xx} = \partial^2 u/\partial x^2$, $u_{xy} = \partial^2 u/\partial x \partial y$, and $u_{yy} = \partial^2 u/\partial y^2$ are the derivatives of u; and $F(x,y,u, p,q,r,s,t)$ is a given function of eight variables. For example, Eq. (4) is a partial differential equation of

$$u_{xy} = 0 \tag{4}$$

order 2. It is easily verified that $u = h(x) + g(y)$ is a solution of Eq. (4) for any choice of functions h and g.

A differential equation is linear if the function u and its derivatives appear linearly in the equation, and is nonlinear otherwise. Thus, Eqs. (2) and (4) are linear, whereas Eqs. (5), (6), and (7) are nonlinear.

$$u' = \sin u \tag{5}$$

$$(1 + u_y^2)\, u_{xx} - 2u_x u_y u_{xy} + (1 + u_x^2) u_{yy} = 0 \tag{6}$$

$$u_t + uu_x + u_{xxx} = 0 \tag{7}$$

Partial differential equations occur with more than two independent variables; see Eqs. (33), (35), and (36). Another possibility is a system of differential equations. This consists of two or more equations involving one or more unknown functions which are to be solved simultaneously. For example, Eqs. (8) are a first-order linear system of ordinary dif-

$$\begin{aligned} u' &= v \\ v' &= -u \end{aligned} \tag{8}$$

ferential equations. The Cauchy-Riemann equations, Eqs. (9), are a first-order linear system of partial dif-

$$\begin{aligned} u_x - v_y &= 0 \\ u_y + v_x &= 0 \end{aligned} \tag{9}$$

ferential equations. *See* COMPLEX NUMBERS AND COMPLEX VARIABLES.

Ordinary differential equations. The process of solving differential equations is called integration, and is analogous to the solution of the simple equation $u'(t) = h(t)$, which is actually the indefinite integral of h. The value of an indefinite integral is unique except for an arbitrary constant of integration. In general, the solution of an ordinary differential equation of order n involves n arbitrary constants of integration. For example, the general solution of Eq. (2) is $u(t) = A \cos t + B \sin t$, where A and B are arbitrary constants; and the solution of Eq. (5) is $u(t) = 2 \tan^{-1} (ke^t)$, where k is an arbitrary constant. *See* INTEGRATION.

In the full generality of Eq. (1), it is not possible to prove that solutions exist. For example, Eq. (10) has

$$1 + \left| \frac{du}{dt} \right|^2 = 0 \tag{10}$$

no solution. However, if an equation is of the type of Eq. (11), and if $f(t,u_0, \ldots, u_{n-1})$ has continuous

$$u^{(n)} = f(t, u, \ldots, u^{(n-1)}) \tag{11}$$

first derivatives on the set defined by $|t - t_0| < d$, $|u_0 - a_0| < d$, $|u_1 - a_1| < d, \ldots, |u_{n-1} - a_{n-1}| < d$, where $a_1, \ldots, a_n$, and d are constants, then there is one and only one function $u(t)$ defined and continuously differentiable on the interval $|t - t_0| < \epsilon$, for some $\epsilon > 0$, which is a solution of Eq. (11) and which satisfies the initial conditions $u(t_0) = a_0$, $u'(t_0) = a_1, \ldots, u^{(n-1)}(t_0) = a_{n-1}$. This result is purely an existence result. That is, it says a solution exists but does not say how to find it. Indeed, it may easily be impossible to find a solution which is expressible in terms of elementary functions.

Equations of first order. It is usual in the case of equations of first order to use x as the independent variable and to denote the unknown function by $y = y(x)$. Then Eq. (1) reduces to $dy/dx = f(x,y)$. This equation is usually written in differential form, Eq. (12), where $f = -P/Q$. In this notation, Eq. (5)

$$P(x, y)dx + Q(x, y)dy = 0 \tag{12}$$

would be rewritten as Eq. (13). For a given func-

$$\sin y dx - dy = 0 \tag{13}$$

tion $v(x,y)$ the differential of v is defined by $dv = v_x dx + v_y dy$. The first-order equation is said to be exact if there is a function v such that $dv = Pdx + Qdy$,

that is, if $\partial v/\partial x = P$ and $\partial v/\partial y = Q$. In this case the solution to Eq. (12) is given implicitly by the equation $v(x,y) = c$, with c an arbitrary constant. For example, Eq. (14) is exact, since $d(xy) = ydx + xdy$, and

$$ydx + xdy = 0 \tag{14}$$

the solution is $xy = c$. A general example of exact equations is given by the case when, in Eq. (12), P is a function of x only and Q is a function of y only. Then, if $v(x,y) = \int P(x)dx + \int Q(y)\,dy$, the solution is $v(x,y) = c$. For example, Eq. (15) is exact and has

$$dx - \frac{dy}{\sin y} = 0 \tag{15}$$

the solution $x - \log \tan (y/2) = c$.

When Eq. (12) is not exact, it may be possible to find an integrating factor, that is, a function $\mu(x,y)$ such that the equation $\mu Pdx + \mu Qdy$ is exact. If $\mu Pdx + \mu Qdy = dv$, then, once more, the solution is given by $v(x,y) = c$. Theoretically it is always possible to find an integrating factor, provided P and Q are not both equal to zero at some point. In practice, however, it may be difficult. For Eq. (16), which is not exact, an integrating factor is

$$ydx - xdy = 0 \tag{16}$$

$1/y^2$, since $d(x/y) = (ydx - xdy)/y^2$, and the solution is given by $x/y + c$. A general example is provided by the case of separable variables when, in Eq. (12), $P(x,y) = A(x)B(y)$ and $Q(x,y) = C(x)D(y)$. Then $\mu = 1/C(x)B(y)$ is an integrating factor, since $[A(x)/C(x)]\,dx + [D(y)/B(y)]\,dy = 0$ is exact. Using this method, Eq. (13), and hence Eq. (5), can be reduced to Eq. (15). As another example, consider Eq. (17), a linear first-order differential equation, or in differential form, Eq. (18). It is not exact, but $\mu(x) =$

$$\frac{dy}{dx} + a(x)y = z(x) \tag{17}$$

$$dy + [a(x)y - z(x)]dx = 0 \tag{18}$$

$\exp\left[\int a(x)dx\right]$ is an integrating factor. After multiplication by μ, Eq. (18) becomes $dv = 0$, where

$$v(x,y) = y\exp\left[\int a(x)dx\right] - \int z(x)\exp\left[\int a(x)dx\right]dx$$

Thus the general solution of Eq. (17) is given by Eq. (19), where c is an arbitrary constant.

$$y(x) = \exp\left[-\int a(x)dx\right] \cdot \left\{\int z(x)\exp\left[\int a(x)dx\right]dx + c\right\} \tag{19}$$

Second-order linear equations. The study of oscillation in mechanical systems and of simple electric circuits leads to linear ordinary differential equations of second order, typified by Eq. (20). This equation is said

$$u'' + a(t)u' + b(t)u = v \tag{20}$$

to be homogeneous when v is identically equal to zero; otherwise, it is nonhomogeneous. The linearity of the equation has important implications for the sets of solutions to Eq. (20). If u_1 and u_2 are two linearly independent solutions to the homogeneous equation, then $Au_1 + Bu_2$ is also a solution for any choice of constants A and B. Furthermore, any solution of the homogeneous equation is of this form. Since the difference of two solutions to the nonhomogeneous equation is a solution to the homogeneous equation, it follows that the general solution to the nonhomogeneous equation has the form of expression (21), where u_p is a particular solution to the

$$u_p + Au_1 + Bu_2 \tag{21}$$

nonhomogeneous equation, u_1 and u_2 are linearly independent solutions to the homogeneous equation, and A and B are arbitrary constants. As an example, consider Eq. (22). The corresponding homogeneous

$$u'' + u = t \tag{22}$$

equation is Eq. (2). The functions $u_1(t) = \cos t$ and $u_2(t) = \sin t$ are linearly independent solutions to Eq. (2); thus the general solution to Eq. (2) is $u(t) = A\cos t + B\sin t$. It is easy to find a particular solution to Eq. (22), namely, $u_p(t) = t$. Hence the general solution to Eq. (22) is $u(t) = t + A\cos t + B\sin t$.

When the coefficients $a(t)$ and $b(t)$ in Eq. (20) are constants, solutions to the homogeneous equation exist of the form $u(t) = e^{rt}$, where r is a constant which may be complex. Substitution into Eq. (20) [with $v = 0$] yields the quadratic equation $r^2 + ar + b = 0$, which may be solved for r. If $a^2 - 4b > 0$, there are two distinct real roots, and thus two linearly independent solutions to the homogeneous equation exist. If $a^2 - 4b = 0$, there is only one root. A second solution in this case is $u(t) = te^{rt}$. If $a^2 - 4b < 0$, the roots are complex conjugate, $r_1 = \alpha + i\beta$, $r_2 = \alpha - i\beta$. Then two real linearly independent solutions to the homogeneous equation are $e^{\alpha t}\cos \beta t$ and $e^{\alpha t}\sin \beta t$. *See* DAMPING; ELECTRIC TRANSIENT; FORCED OSCILLATION; HARMONIC MOTION.

If the coefficients of Eq. (20) are nonconstant, it is sometimes difficult to find solutions to the homogeneous equation. However, if two linearly independent solutions u_1 and u_2 to the homogeneous equation are known, the method of variation of parameters yields solutions to the nonhomogeneous equation. The idea is to look for a solution to Eq. (20) of the form $A(t)u_1(t) + B(t)u_2(t)$, subject to the condition that $A'u_1 + B'u_2 = 0$. After solving for A and B, formula (23) is obtained for a particular solution to Eq. (20), where the kernel $k(t,s)$ is given by Eq. (24).

$$u_p(t) = \int_{t_0}^{t} k(t,s)v(s)ds \tag{23}$$

$$k(t,s) = \frac{u_2(t)u_1(s) - u_1(t)u_2(s)}{u_1(s)u_2'(s) - u_2(s)u_1'(s)} \tag{24}$$

If the coefficients $a(t)$ and $b(t)$ of Eq. (20) are

analytic functions (that is, if a and b have power series expansions

$$a(t) = \sum_{n=0}^{\infty} a_n t^n, \quad b(t) = \sum_{n=0}^{\infty} b_n t^n$$

which converge for t near zero), then power series solutions to the homogeneous equation can be found. The idea is to set

$$u(t) = \sum_{n=0}^{\infty} u_n t^n$$

Then

$$u'(t) = \sum_{n=0}^{\infty} n u_n t^{n-1} \text{ and } u''(t) = \sum_{n=0}^{\infty} n(n-1) u_n t^{n-2}$$

The series for a, b, u, u', and u'' are substituted into the homogeneous equation, resulting in a power series which is identically equal to zero. Setting the coefficients equal to zero results in a sequence of recursion relations for the coefficients u_n. This method can be extended to the case for which not $a(t)$ and $b(t)$ themselves but $ta(t)$ and $t^2b(t)$ are analytic functions. This extension is essential since many important equations have this feature. For example, the solution of Bessel's equation, Eq. (25), is Bessel's function of order n, shown in Eq. (26), where Γ de-

$$t^2u'' + tu' + (t^2 - n^2)u = 0 \tag{25}$$

$$J_n(t) = \left(\frac{t}{2}\right)^n \sum_{k=0}^{\infty} \frac{(-1)^k (t/2)^{2k}}{\Gamma(k+1)\Gamma(n+k+1)} \tag{26}$$

notes the gamma function of Euler. Another example is the hypergeometric equation of Gauss, Eq. (27). If

$$t(1-t)u'' + [c - (a+b+1)t]u' - abu = 0 \tag{27}$$

c is not zero or a negative integer, a solution is the hypergeometric function shown in Eq. (28).

$$F(a,b,c,t) = 1 + \frac{a \cdot b}{1 \cdot c} t + \frac{a(a+1)b(b+1)}{1 \cdot 2 \cdot c(c+1)} t^2 + \cdots \tag{28}$$

See BESSEL FUNCTIONS; GAMMA FUNCTION.

Partial differential equations. The analysis of many physical problems leads to linear partial differential equations of second order. When there are two independent variables x and y, the most general equation of this sort is Eq. (29), where the coefficients $a, \ldots,$

$$au_{xx} + 2bu_{xy} + cu_{xx} + du_x + eu_y + fu = g \tag{29}$$

g are given functions of x and y. It will be assumed that $a^2 + b^2 + c^2 \neq 0$. Then Eq. (29) is called elliptic if $b^2 - ac < 0$, parabolic if $b^2 - ac = 0$, and hyperbolic if $b^2 - ac > 0$. For each of these types there are certain typical boundary value problems which arise in physical applications and which are of mathematical interest. When there are more than two independent variables, or when the order of the equation is larger than two, there are equations which have the same qualitative properties, as do these three types, and for which similar boundary value problems exist. These equations are accordingly called elliptic, parabolic, or hyperbolic. However, some equations do not belong to any of these three categories.

Elliptic equations. The prototype is Laplace's equation, Eq. (30), for two independent variables, and Eq. (31) for three independent variables. The solu-

$$u_{xx} + u_{yy} = 0 \tag{30}$$

$$u_{xx} + u_{yy} + u_{zz} = 0 \tag{31}$$

tions are called harmonic functions. A typical boundary value problem is the Dirichlet problem. It consists of finding a harmonic function $u(x,y)$ in a region in the plane [say, the disk $D = \{(x,y) | x^2 + y^2 < 1\}$] with prescribed values on the boundary of the region [$u(x,y) = g(x,y)$ for $x^2 + y^2 = 1$, where g is a given function]. This corresponds to the physical problem of finding the steady-state temperature distribution $u(x,y)$ in a plate when the temperature distribution on the edge of the plate is known. *See* LAPLACE'S DIFFERENTIAL EQUATION.

Parabolic equations. The prototype is the heat equation, shown in one space variable x in Eq. (32), and in two space variables x and y in Eq. (33). The vari-

$$\frac{\partial u}{\partial t} - \frac{\partial^2 u}{\partial x^2} = 0 \tag{32}$$

$$\frac{\partial u}{\partial t} - \frac{\partial^2 u}{\partial x^2} - \frac{\partial^2 u}{\partial y^2} = 0 \tag{33}$$

able t represents time. If $u(t,x)$ is the temperature in a rod modeled by the interval $a \leq x \leq b$, and if there is no source of heat in the rod, then u is a solution to Eq. (32). A typical boundary value problem is the initial value problem, which consists of finding a solution to Eq. (32) for $0 < t < \infty$ and $-\infty < x < \infty$ with $u(0,x) = f(x)$ given. The physical interpretation consists of finding the temperature distribution in an infinite rod for all time when the initial temperature distribution is known. *See* CONDUCTION (HEAT).

Hyperbolic equations. The prototype is the wave equation, shown for one, two, and three space variables in Eqs. (34), (35), and (36), respectively. Again, t rep-

$$\frac{\partial u}{\partial t} - \frac{\partial^2 u}{\partial x^2} - \frac{\partial^2 u}{\partial y^2} = 0 \tag{34}$$

$$\frac{\partial^2 u}{\partial t^2} - \frac{\partial^2 u}{\partial x^2} - \frac{\partial^2 u}{\partial y^2} = 0 \tag{35}$$

$$\frac{\partial^2 u}{\partial t^2} - \frac{\partial^2 u}{\partial x^2} - \frac{\partial^2 u}{\partial y^2} - \frac{\partial^2 u}{\partial z^2} = 0 \tag{36}$$

resents time. Equation (34) is also called the equation of the vibrating string; Eq. (35) is called the equation of the vibrating membrane. As these names indicate, a large variety of oscillatory phenomena lead to hyperbolic equations. A typical boundary value problem is the Cauchy problem, which involves finding a solution $u(t, x)$ to Eq. (34) for $0 < t < \infty$ and $-\infty$

$< x < \infty$, with $u(0, x) = f(x)$ and $u_t(0, x) = g(x)$ given. If $u(t, x)$ is interpreted as the displacement of an infinite string, then the Cauchy problem consists of deriving $u(t, x)$ for all time when the initial displacement $u(0,x) = f(x)$ and the initial velocity $u_t(0, x) = g(x)$ are both known. *See* VIBRATION; WAVE EQUATION; WAVE MOTION.

Other equations. An example of a partial differential equation which is not of the types previously considered is the Korteweg-deVries (KdV) equation, Eq. (7), which describes shallow-water waves in a one-dimensional channel. The KdV equation is a nonlinear equation of the third order. Another nonlinear equation of interest is Eq. (6), which is called the minimal surface equation; its solution represents surfaces of minimal surface area (for example, soap films spanning wire frames have minimal surface area). *See* CALCULUS; DIFFERENTIATION. John C. Polking

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Differential geometry

A branch of mathematics that deals with intrinsic properties of curves and surfaces in three-dimensional euclidean space. The intrinsic properties are those which are independent of the geometrical object's orientation or location in space. The subject is also concerned with nets of curves and families of surfaces, these having wide application in the arts.

The space is referred to a rectangular cartesian coordinate system (x, y, z). A space curve may be defined by a pair of independent equations $f(x, y, z) = 0$ and $g(x, y, z) = 0$ or, more meaningfully, by parametric equations: $x = x(t), y = y(t), z = z(t)$. In this case the arc length between two points t_0 and t is given by Eq. (1).

$$s = \int_{t_0}^{t} \sqrt{\left(\frac{dx}{dt}\right)^2 + \left(\frac{dv}{dt}\right)^2 + \left(\frac{dz}{dt}\right)^2}\, dt \qquad (1)$$

Obviously, if s is chosen as a parameter, Eq. (2) holds. In this case dx/ds, dy/ds, dz/ds at a

$$\left(\frac{dx}{ds}\right)^2 + \left(\frac{dy}{ds}\right)^2 + \left(\frac{dz}{ds}\right)^2 = 1 \qquad (2)$$

point P are the direction cosines of the tangent to the curve at P.

Plane curves. Consider three nearby points P, P_1, P_2. Through them, in general, one plane may be constructed. The limiting position of this plane as P_1 and P_2 approach P is the osculating plane of the curve at P. The limiting position of the circle through P, P_1, P_2 as P_1 and P_2 approach P is the osculating circle. Its center is the center of curvature and its radius is the radius of curvature ρ. The reciprocal of the radius of curvature κ is the curvature; its value is shown in expression (3). The perpendicular to the tangent at

$$\sqrt{\left(\frac{d^2x}{ds^2}\right)^2 + \left(\frac{d^2y}{ds^2}\right)^2 + \left(\frac{d^2z}{ds^2}\right)^2} \qquad (3)$$

P in the osculating plane is the principal normal and the perpendicular at P to the osculating plane is the binormal. Another way of defining curvature is this: Let P' be a point near P, arc $PP' = \Delta s$. Let $\Delta\theta$ be the angle between the tangents at P and P'. Then Eq. (4) holds.

$$\frac{1}{\rho} = \lim_{P' \to P} \frac{\Delta\theta}{\Delta s} \qquad (4)$$

Similarly, if $\Delta\phi$ is the angle between the binormals at P and P', expression (5) is the torsion $1/\tau$ of the

$$\lim_{P' \to P} \frac{\Delta\phi}{\Delta s} \qquad (5)$$

curve at P, where τ is the radius of torsion. The torsion at P is the rate at which the curve is twisted out of the osculating plane so that the torsion of a plane curve is zero. The values of the curvature and torsion in terms of the arc determine the curve up to its position in space. The determination of the parametric equations of the curve, as well as of the solution of most problems dealing with space curves, is accomplished by means of the Frenet equations. These equations give the values of the derivatives of the direction cosines of the tangent, normal, and binormal in terms of these cosines and the curvature and torsion of the curve.

Plane curves are, of course, special cases of space curves ($\tau = 0$). One of the problems treated extensively in classical differential geometry is that of nets of curves. Such a net is given by two independent equations, Eqs. (6). For fixed values of υ, Eqs. (6)

$$x = x(u, \upsilon) \qquad y = y(u, \upsilon) \qquad (6)$$

define a family of curves, the u lines, whereas fixed values of u give the υ lines. Through each point there is one u line and one υ line so that (u,υ) may be regarded as curvilinear coordinates in the plane.

If this system is to be orthogonal, it is necessary that Eq. (7) hold.

$$\frac{\partial x}{\partial u}\frac{\partial x}{\partial v} + \frac{\partial y}{\partial u}\frac{\partial y}{\partial v} = 0 \qquad (7)$$

Equations (6) may also be interpreted as the mapping of the $u\upsilon$ plane on the xy plane. In applications one seeks a mapping under which certain properties of figures are preserved. If arc length is to be preserved, the mapping must be a motion; if area is to be preserved, it is necessary that Eq. (8) hold.

$$\frac{\partial x}{\partial u}\frac{\partial y}{\partial v} - \frac{\partial y}{\partial u}\frac{\partial x}{\partial v} = 1 \qquad (8)$$

Probably the most important mappings are the conformal ones. Under them the angles between lines are preserved; such a mapping is given by $x + iy = f(u + i\upsilon)$, where f is any analytic function of the complex variable $u + i\upsilon$ ($i = -1$).

Surfaces. A surface in three-dimensional euclidean space may be given by $f(x, y, z) = 0$ or, more conveniently, by Eqs. (9). For a fixed value of υ, Eqs. (9)

$$x = x(u, v) \qquad y = y(u, v) \qquad z = z(u, v) \tag{9}$$

describe a curve in the surface, a u line, and similarly for a fixed value of u. Thus (u,υ) are curvilinear coordinates of a point in the surface. The most important quantity in the study of surfaces is the arc length of a curve. This is given by Eq. (10), where E, F, $G$are

$$ds^2 = E\,du^2 + 2F\,du\,dv + G\,dv^2 \tag{10}$$

functions of partial derivatives of (x, y, z). In the real domain E, G, and $EG - F^2$ are nonnegative and may be zero only at singular points of the surface, or at points where the matrix of the partial derivatives of (x, y, z) is of rank less than two. The right-hand side of Eq. (10) is called the first fundamental form of the surface, where E, F, G are the first fundamental quantities of the surface. All applicable surfaces have the same first fundamental form. Thus any cylinder or cone which is developable (applicable to a plane) has the fundamental form $du^2 + d\upsilon^2$, where (u, υ) are rectangular cartesian coordinates. Under an arbitrary transformation of the surface coordinates (u, υ) the fundamental quantities E, F, G transform linearly so that many problems in the study of surfaces reduce to the question of whether a coordinate system exists on the surface in which E, F, G satisfy desired conditions.

In order to distinguish between various surfaces having the same first fundamental form, one considers the plane tangent to the surface at a point M of coordinates (u, υ) and the perpendicular distance p from the point M' of coordinates $(u + du, \upsilon + d\upsilon)$ to this tangent plane. The principal part of $2p$ is then expression (11), where D, D', D'' are functions of

$$D\,du^2 + 2D'\,du\,dv + D''\,dv^2 \tag{11}$$

the first and second partial derivatives of (x, y, z) and are therefore expressible as functions of u and v. Expression (11) is called the second fundamental form of the surface. E, F, G and D, D', D'' are not independent but must satisfy the so-called Gauss-Codazzi equations. If both forms are given, the surface is determined up to its position in space.

At each nonsingular point P of a surface, there is one tangent plane and one surface normal. A plane in a given direction containing this normal cuts out a normal section of the surface. The curvature of this curve at P is the normal curvature of the surface in the direction of the plane. As the normal plane rotates about the surface normal as an axis, the normal curvature usually changes; it takes on its maximum and its minimum values for two directions, and these are at right angles to each other. These are the principal directions at P. If all normal curvatures at P have the same value, P is called an umbilical point. Thus every point on a sphere is umbilical, and a general ellipsoid has four umbilical points. The curves on the surface whose tangents have a principal direction at every point are the lines of curvature of the surface. They are often chosen as the coordinate lines, for in that case $F = 0$, $D' = 0$. Denoting by ρ_1 and ρ_2 the radii of curvature of the principal normal sections at P, Eq. (12) is called the mean curvature of the surface

$$K_m = \frac{1}{\rho_1} + \frac{1}{\rho_2} \tag{12}$$

at P and $K = 1/\rho_1\rho_2$ is the total or gaussian curvature of the surface at P. The value of K is given by Eq. (13). Asthus defined, K_m and K depend on the

$$K = \frac{DD'' - D'^2}{EG - F^2} \tag{13}$$

enveloping space. However K. F. Gauss showed that K is expressible in terms of E, F, G and their first and second partial derivatives. Thus K is an intrinsic invariant of the surface; it has the same value for all applicable surfaces. For a plane $K = 0$, for a sphere $K = 1/a^2$, and for a pseudosphere $K = -(1/a^2)$.

Through each point of a surface in a given direction there is one curve whose osculating plane coincides with the normal plane in that direction. These curves are the geodesics of the surface. In a properly restricted neighborhood of a point there is just one geodesic through two points in that neighborhood; it is the shortest distance between the two points. From this it follows that the geodesics are tangent to the curve at the points s and $s + \Delta s$. This curvature κ_g is measured in the surface and is in general different from the curvature of the curve as measured in space. Obviously κ_g is zero for any geodesic, so that geodesics play the role of straight lines in the plane.

n-Dimensional applications. The idea of obtaining all intrinsic metric properties of a surface from the first fundamental form was extended by G. F. B. Riemann to a space of n dimensions. In this type of space a point has the coordinates $(u_1, u_2, \ldots, u_n)$ and the arc length is given by Eq. (14). Such a space is called

$$ds^2 = \sum_{i,j=1}^{n} g_{ij}(u)du_i\,du_j \tag{14}$$

riemannian and the functions $g_{ij}(u)$ are the fundamental quantities for the space. All the intrinsic properties of a surface have been generalized to a riemannian space though not in a trivial or obvious manner. It was the knowledge of this geometry that enabled Albert Einstein to work out the general theory of relativity, for according to that theory, physical space is four dimensional, the $g_{ij}(u)$ being determined by the matter and energy in the space. *See* CONFORMAL MAPPING; COORDINATE SYSTEMS; PROJECTIVE GEOMETRY; RELATIVITY; RIEMANNIAN GEOMETRY; TENSOR ANALYSIS.

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Differential transformer

An iron-core transformer with movable core. A differential transformer produces an electrical output voltage proportional to the displacement of the core. It is used to measure motion and to sense displacements. It is also used in measuring devices for force, pressure, and acceleration that are based on the conversion of the measured variable to a displacement.

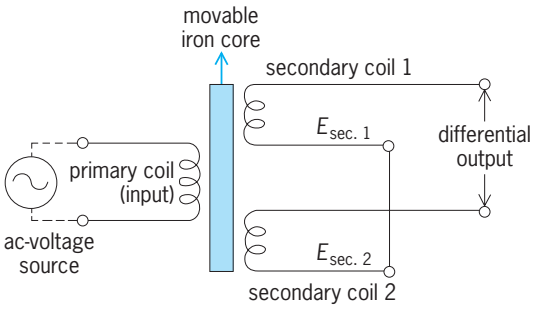

Fig. 1. Basic circuit of differential transformer.

Various available configurations, some translational and others rotational, all employ the basic circuit of **Fig. 1**: a primary winding, two secondary windings, and a movable core. The primary winding is energized with alternating voltage. The two secondary windings are connected in series opposition, so that the transformer output is the difference of the two secondary voltages. When the core is centered, the two secondary voltages are equal and the transformer output is zero. This is the balance of null position. When the core is displaced from the null point, the two secondary voltages are no longer alike and the transformer produces an output voltage. With proper design, the output voltage varies linearly with core position over a small range. Motion of the core in the opposite direction produces a similar effect with 180° phase reversal of the alternating output voltage.

The amplitude of the ac output voltage forms a V-shaped curve when plotted against core position; the phase angle abruptly reverses by 180° at the null point. When the bottom of the V-curve is examined in closer detail, it is seen that the output voltage at balance is not exactly zero. The small residual null voltage consists of higher harmonics of the input frequency, as well as a fundamental frequency component 90 degrees out of phase (called the quadrature component).

Electronic signal conditioning is commonly employed to eliminate the residual and to make the transducer usable with standard dc instrumentation.

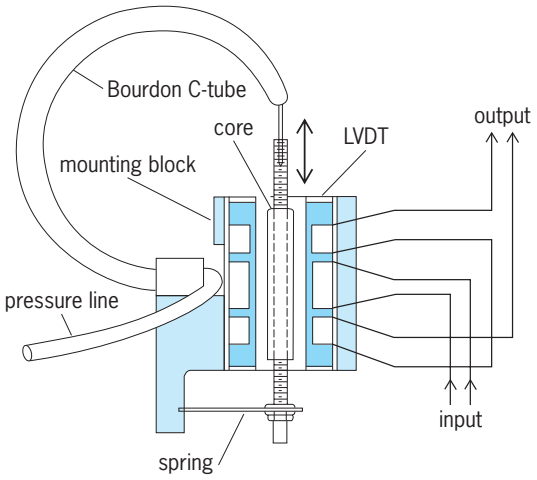

Fig. 2. Bourdon-tube pressure transducer employing a linear variable differential transformer (LVDT). (*After E. E. Herceg, Handbook of Measurement and Control, Schaevitz Engineering, 1976*)

The electronic circuit consists of an ac oscillator (carrier generator) to drive the input windings, plus a demodulator and an amplifier to convert the output into dc. The microelectronics can be built into the transformer housing, and the resulting package is sold as a dc-LVDT.

The principal advantages of the differential transformer, which are shared with capacitance transducers, over other displacement transducers, such as the resistance potentiometer, are absence of contacts and infinite resolution. No friction is introduced by the measurement, and movement smaller than a microinch (25 nanometers) can be sensed. The separation between coil and core makes the differential transformer useful in difficult and dangerous environments. Stability of the null makes it ideal as a null sensor in self-balancing devices and servomechanisms. Typical applications are machine tool inspection and gaging, pressure measurement (**Fig. 2**), liquid level control, load cells, and gyroscopic instruments.

The linear variable differential transformer (LVDT; **Fig. 3**) is the commercially prevalent form. A rotary version is also manufactured. The E-pickoff (**Fig. 4**) is an older device. Its linearity is not as good, and its principal use is as a null sensor. Both translational

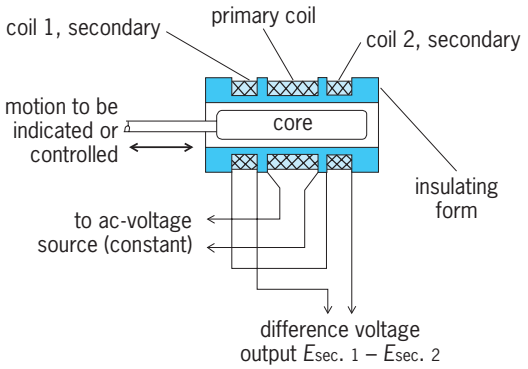

Fig. 3. Linear variable differential transformer (LVDT).

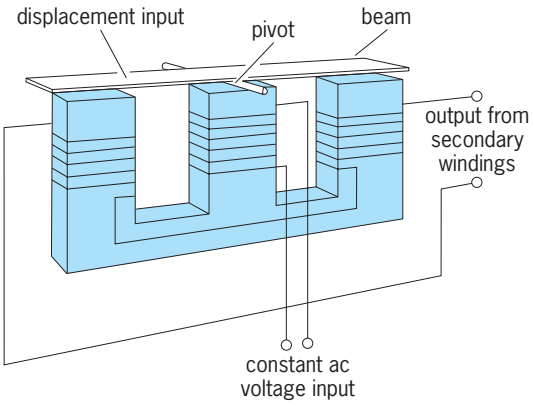

Fig. 4. E-shaped differential transformer. (*After P. J. O'Higgins, Basic Instrumentation, McGraw-Hill, 1966*)

and rotational E-pickoffs are made. *See* TRANSDUCER; TRANSFORMER. Gerald Weiss

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Differentiation

A mathematical operation performed on a function to determine the effect of a change in the value of the independent variable. Functions of one variable are considered in this article. For differentiation of functions of several variables *see* PARTIAL DIFFERENTIATION.

If f is a function of x, defined on an interval containing x_0, the derivative at x_0 is, by definition, that shown in Eq. (1). For generalities about derivatives

$$f'(x_0) = \lim_{x \to x_0} \frac{f(x) - f(x_0)}{x - x_0} \tag{1}$$

and calculus *see* CALCULUS.

In the quotient on the right, in the definition of $f'(x_0)$, x is restricted to the interval on which f is defined. If $y = f(x)$, $f'(x_0)$ is also denoted by expression(2). The limit defining $f'(x_0)$ may not exist. If it

$$\left(\frac{dy}{dx}\right)_{x=x_0} \tag{2}$$

does, f is called differentiable at x_0.

The derivative of $f'(x)$, called the second derivative, is denoted by $f''(x)$ or d^2y/dx^2.

A function f is called continuous at x_0 if x_0 is in the domain of f and if Eq. (3) holds. The precise

$$\lim_{x \to x_0} f(x) = f(x_0) \tag{3}$$

formulation of the limit concept used here is the following: Let g be a function and let A be a number. Then Eq. (4) means that to each positive number

$$\lim_{x \to x_0} g(x) = A \tag{4}$$

ϵ corresponds to some positive number δ such that $|g(x) - A| < \epsilon$ whenever x is a number in the domain of g such that $x \neq x_0$ and $| x - x_0 | < \delta$. It is required of g that its domain shall contain numbers x as close to x_0 as desired, but different from x_0. The domain of g may also contain x_0, but this is irrelevant.

It is a theorem that, if f is differentiable at x_0, then f is continuous at x_0. However, a function can be continuous at x_0, but not differentiable there. An example is $f(x) = |x|$ at $x_0 = 0$. It is even possible for a function to be continuous on an interval and yet not differentiable at any point of this interval.

The chief elementary applications of differentiation are (1) in expressing rates of change (velocity, acceleration), and in solving problems where, through functional relationship, the rate of change of one variable is calculated when the rate of change of another variable is known; (2) in studying graphs of functions and, more generally, in studying curves in the plane or in space of three dimensions; (3) in expressing scientific laws or principles in the form of differential equations; and (4) in the expression of various extensions and applications of the law of the mean, including such topics as l'Hospital's rule and Taylor's formula or series.

Principles. The general technique of differentiation is built upon the rules for differentiating combinations of differentiable functions. If $u = f(x)$ and $v = g(x)$ are functions differentiable for the same values of x, then $u + v$ and uv are differentiable and so is u/v if $v \neq 0$. The formulas are shown in Eqs. (5). If

$$\begin{aligned} \frac{d}{dx}(u+v) &= \frac{du}{dx} + \frac{dv}{dx} \\ \frac{d}{dx}(uv) &= u\frac{dv}{dx} + v\frac{du}{dx} \\ \frac{d}{dx}\left(\frac{u}{v}\right) &= \frac{v\dfrac{du}{dx} - u\dfrac{dv}{dx}}{v^2} \end{aligned} \tag{5}$$

u is constant in value, then $du/dx = 0$. A very powerful instrument of technique is furnished in the chain rule for composite functions. If y is a differentiable function of u and u is a differentiable function of x, then y is a differentiable function of x, and Eq. (6)

$$\frac{dy}{dx} = \frac{dy}{du}\frac{du}{dx} \tag{6}$$

results. In functional notation, if $y = f(u)$ and $u = h(x)$, then $y = F(x)$, where $F(x) = f[h(x)]$, and then $f'(x) = f'[h(x)]h'(x)$.

The technique of differentiation also leans upon facts about inverse functions. If $y = f(x)$, where f is a differentiable function and $f'(x)$ is either always positive or always negative on an interval, for example, when $a \leqq x \leqq b$, then for each y from $f(a)$ to $f(b)$ inclusive there is just one x such that $a \leqq x \leqq b$ and $y = f(x)$. Thus there is defined a function g such that $x = g(y)$ is equivalent to $y = f(x)$ with x and y restricted as indicated. This function g is differentiable, and Eq. (7) results. An example: $y = \sin x$,

$$g'(y) = \frac{1}{f'(x)} = \frac{1}{f'[g(y)]} \tag{7}$$

$x = \sin^{-1} y$ (sometimes written $x = \arcsin y$), where $-\pi/2 \leqq x \leqq \pi/2$ and $-1 \leqq y \leqq 1$. Here $f'(x) = \cos x$, and Eq. (8) gives the derivative of the inverse function.

$$\frac{d}{dy}(\sin^{-1} y) = \frac{1}{\cos x} = \frac{1}{\sqrt{1 - \sin^2 x}} = \frac{1}{\sqrt{1 - y^2}} \tag{8}$$

Algebraic and transcendental functions. The functions studied in elementary calculus are of two kinds, algebraic and transcendental. The basic differentiation formula for algebraic functions is $dy/dx = nx^{n-1}$ if $y = x^n$, where n is any rational number. This rule may be combined with the chain rule and used in connection with the rules for dealing with sums, products, and quotients. The differentiation of algebraic functions in general may require use of implicit function theorems. The elementary transcendental functions are the trigonometric functions, the logarithm functions, and their inverses. For a discussion of implicit function theorems *see* PARTIAL DIFFERENTIATION.

Trigonometric functions. In calculus the trigonometric functions are defined on the assumption that angles are measured in radians, so that $\sin x$ means the sine of x radians. Differentiation of $f(x) = \sin x$ is based on the fact that $\sin x/x \to 1$ as $x \to 0$. This is equivalent to $f'(0) = 1$. By combining this result with trigonometric identities, the two basic formulas shown in Eqs. (9) are derived. Then the derivatives of the other

$$\frac{d}{dx}\sin x = \cos x \qquad \frac{d}{dx}\cos x = -\sin x \tag{9}$$

functions are worked out by using the rule for quotients. The results are shown in Eqs. (10). These formulas may also be combined with the chain rule; an example is given in Eq. (11).

$$\begin{aligned} \frac{d}{dx}\tan x &= \sec^2 x \frac{d}{dx}\cot x = -\csc^2 x \\ \frac{d}{dx}\sec x &= \sec x \tan x \frac{d}{dx}\csc x = -\csc x \cot x \end{aligned} \tag{10}$$

$$\frac{d}{dx}\sin u = \cos u \frac{du}{dx} \tag{11}$$

The inverse trigonometric functions are differentiated by the methods for inverse functions, as explained earlier. The definitions are as follows:

$y = \sin^{-1} x$ means $x = \sin y$ and $-\dfrac{\pi}{2} \leqq y \leqq \dfrac{\pi}{2}$

$y = \cos^{-1} x$ means $x = \cos y$ and $0 \leqq y \leqq \pi$

$y = \tan^{-1} x$ means $x = \tan y$ and $-\dfrac{\pi}{2} < y < \dfrac{\pi}{2}$

$y = \cot^{-1} x$ means $x = \cot y$ and $0 < y < \pi$

The differentiation formulas are shown in Eqs. (12).

$$\begin{aligned} \frac{d}{dx}\sin^{-1} x &= \frac{1}{\sqrt{1 - x^2}} \\ \frac{d}{dx}\cos^{-1} x &= \frac{-1}{\sqrt{1 - x^2}} \\ \frac{d}{dx}\tan^{-1} x &= \frac{1}{1 + x^2} \\ \frac{d}{dx}\operatorname{ctn}^{-1} x &= \frac{-1}{1 + x^2} \end{aligned} \tag{12}$$

The inverses of the secant and cosecant functions are little used, and there is no standard usage about the definitions needed to make them single-valued.

Exponentials and logarithms. These two functions go together, a logarithm function being the inverse of an exponential function, or vice versa. The traditional treatment of these functions in differential calculus was for many years as follows: The nature of exponentials was assumed as known from algebra, and the definition $y = \log_a x$ if $x = a_y$ (where $a > 0$, $a \neq 1$, and $x > 0$), as well as the algebraic properties of logarithms, was also assumed as known. As a first step toward the differentiation of a logarithm function, it was traditional to show that $(1 + t)^{1/t}$ approaches a limit, denoted by e, as t approaches 0. Moreover, $2 < e < 3$, and $e = 2.718$, approximately. It can then be shown that Eq. (13) is valid. This formula suggests the advantage of choosing $a = e$ as the base for logarithms. The base e logarithm of x, called the natural logarithm of x, is denoted by $\log x$ with no subscript, or by $\ln x$. (The latter notation is favored by engineers and many physical scientists.) Then $y = \ln x$ is equivalent to $x = e^y$, and $y = e^x$ is equivalent to $x = \ln y$. The derivative of e^x can be worked out by the rule for inverse functions. The simple formulas are given in Eqs. (14). For an arbitrary base a, where $a \neq 1$ and $0 < a$, the formulas can be written as Eqs. (15). It is often convenient to know that $\log_a x = (\log_a b)(\log_b x)$. In particular, with $a = 1$, one obtains $\log_a b = (\log_b a)^{-1}$.

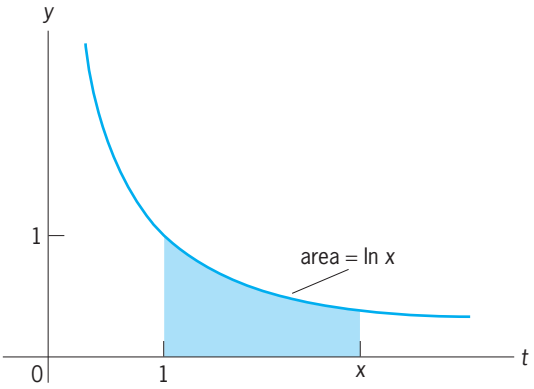
$$\frac{d}{dx}\log_a x = \frac{1}{x}\log_a e \tag{13}$$

$$\frac{d}{dx}\ln x = \frac{1}{x} \qquad \frac{d}{dx}e^x = e^x \tag{14}$$

$$\frac{d}{dx}\log_a x = \frac{\log_a e}{x} \qquad \frac{d}{dx}a^x = a^x \log_e a \tag{15}$$

The modern trend is to recognize that neither exponentials nor logarithms are adequately defined and studied by students before they come to calculus. In any case, satisfactory discussion of these functions requires the theory of limits. Hence it is reasonable to use calculus itself to define logarithms and develop their properties. When this is done, the development proceeds as follows: The "natural logarithm function" is defined when $x > 0$ by Eq. (16).

$$\ln x = \int_1^x \frac{dt}{t} \tag{16}$$



The natural logarithm function.

This means that ln x is the area between the curve $y = 1/t$ and the t axis from $t = 1$ to $t = x$, reckoned as positive if $x > 1$ and as negative if $0 < x < 1$ (see **illus.**).

By use of the fundamental properties of integrals and the relation between differentiation and integration, it can be shown quite directly that $\ln 1 = 0$, $d/dx \ln x = 1/x$, and that $\ln (AB) = \ln A + \ln B$. Moreover, ln x is a continuous function of x which increases as x increases and is such that $\ln x \to +\infty$ as $x \to +\infty$ and $\ln x \to -\infty$ as $x \to 0$. There is then a well-defined inverse function (call it the E function) such that $y = \ln x$ is equivalent to $x = E(y)$ if $x > 0$. The number $E(1)$ is denoted by e; that is, e is the unique positive number such that $\ln e = 1$. The rules about inverse functions show that $E'(x) = E(x)$. If $a > 0$ and x is an integer, it is possible to show easily that $a^x = E(x \ln a)$. This formula serves as the general definition of a^x when x is not an integer, and the results are consistent with what is expected, so that the usual exponent laws are valid. The general definition of logarithms, from this point of view, is Eq. (17). In

$$\log_a x = \frac{\ln x}{\ln a} \tag{17}$$

particular, $\log_e x = \ln x$, because $\ln e = 1$.

In this method of developing properties of exponentials and logarithms by calculus, the fact that $(1 + t)^{1/t} \to e$ as $t \to 0$ is not needed, but it can be proved, more easily than in the older traditional development, after everything else has been worked out.

The number e enters naturally into the concept of continuous compounding of interest. If the sum $\$P$ is placed at interest at the nominal rate of $100r\%$, compounded n times a year, the accumulated sum $\$S$ after t years is shown in Eq. (18). If now n is increased indefinitely, so that interest is compounded more and more frequently, the fact that Eq. (19) holds

$$S = P\left(1 + \frac{r}{n}\right)^{nt} \tag{18}$$

$$\lim_{n\to\infty} \left(1 + \frac{r}{n}\right)^{n/r} = e \tag{19}$$

shows that in the limit of continuously compounded interest, the formula for the accumulated sum after t years is $S = Pe^{rt}$. This formula is characterized by differential equation (20), which expresses what is

$$\frac{dS}{dt} = rS \tag{20}$$

sometimes called the law of natural growth. In general, if y depends on x in such a way that $dy/dx = ky$, where k is a nonzero constant, then $y = y_0 e^{kx}$, where $y = y_0$ when $x = 0$. This situation occurs in radioactive decay and in many types of growth and diminution processes in chemistry and natural science.

Applications. Several important applications of differentiation are discussed in this section.

Velocity and acceleration. If a point moves on the x axis, with coordinate x at time t, its velocity is dx/dt and its acceleration is d^2x/dt^2. When the point with coordinates (x, y) moves in the xy plane, its velocity and acceleration are vectors, the x and y components of velocity are dx/dt and dy/dt, respectively, and the corresponding components of acceleration are d^2x/dt^2, d^2y/dt^2. These results are extended naturally for a point (x, y, z) moving in three-dimensional space.

For the plane case, if the point has polar coordinates (r, θ), the velocity and acceleration can be resolved into components in the r and θ directions. The r direction at (r, θ) is the direction directly away from the origin (r increasing and θ constant). The θ direction is 90° from the r direction in the counterclockwise sense. The r and θ components of velocity are, respectively, dr/dt and $r(d\theta/dt)$, The corresponding components of acceleration are expressions (21) and (22). These are useful in discussing the motion of a

$$\frac{d^2r}{dt^2} - r\left(\frac{d\theta}{dt}\right)^2 \tag{21}$$

$$r\,\frac{d^2\theta}{dt^2} + 2\,\frac{dr}{dt}\,\frac{d\theta}{dt} \tag{22}$$

particle under the influence of a central force, as in the case of a single mass moving in the gravitational field of a fixed center of attraction according to the inverse-square law.

Still another useful way of resolving acceleration into components involves the tangential and normal directions to the path. Here one needs to deal with arc length and curvature. If s is arc length measured along a plane curve in a preassigned positive sense and if K is curvature, then the velocity is a vector whose component along the curve is ds/dt, the component at right angles to the curve being zero. The component of acceleration along the curve in the positive sense is d^2s/dt^2, and at right angles to the curve (90° counterclockwise from the positive sense along the curve) is v^2K, where $v = ds/dt$. The curvature of a circle is the reciprocal of its radius R, and so a point moving around a circle of radius R with speed v experiences an acceleration toward the center of amount v^2/R. This is called centripetal acceleration. If the speed is not uniform, there is also an acceleration along the curve. *See* PARAMETRIC EQUATION.

Curve tracing. If $f'(x) > 0$, $f(x)$ increases as x increases, whereas $f(x)$ decreases as x increases when $f'(x) < 0$. Points where $f'(x) = 0$ are called critical points. With x and y axes in the usual position (x positive to the right, y positive toward the top of the page), the graph of $y = f(x)$ is called concave upward over an interval of x values if the tangent line turns counterclockwise as x increases. If the tangent line turns clockwise as x increases, the curve is called concave downward. If $f'(x) > 0$, the curve is concave upward, and it is concave downward if $f''(x) < 0$. It follows that y is at a relative maximum if $f'(x) = 0$ and $f'(x) < 0$, at a relative minimum if $f'(x) = 0$ and $f'(x) > 0$. A point where the concavity changes from one sense to the other is called a point of inflection. Such a point is a point of relative maximum or minimum for $f'(x)$. A sufficient condition for a point of inflection is that $f''(x) = 0$, $f'''(x) = 0$.

Simple harmonic motion. A point moving on the x axis in such a way that $d^2x/dt^2 = -\omega^2 x$, where $\omega > 0$, is executing what is called simple harmonic motion. If a point travels around a circle with constant speed, its orthogonal projection on a diameter executes simple harmonic motion with the center of the circle as the origin of coordinates on the diameter. The methods of calculus enable one to infer from the foregoing differential equation that x depends on t by the formula $x = A \cos(\omega t + \alpha)$, where A and α are constants. The moving point completes one cycle of its motion in time $2\pi/\omega$.

Law of the mean. For a discussion of the simple case of this law *see* CALCULUS.

An extended version, often called Cauchy's formula, is the following: Suppose F and G are continuous when $a \leqq x \leqq b$, differentiable when $a < x < b$, and suppose $G(b) \neq G(a)$. Suppose also that $f'(x)$ and $G'(x)$ are never both zero together. Then there is some number X such that $a < X < b$ and Eq. (23) is satisfied.

$$\frac{F(b) - F(a)}{G(b) - G(a)} = \frac{F'(X)}{G'(X)} \tag{23}$$

Rule of l'Hospital. This rule for finding the limit of a quotient of two functions in certain circumstances is stated as follows: Subject to certain general conditions of f and g, if either $f(x) \to 0$ and as $x \to a$ or $g(x) \to +\infty$ or $g(x) \to -\infty$ a, then Eq. (24) is valid,

$$\lim_{x \to a} \frac{f(x)}{g(x)} = \lim_{x \to a} \frac{f'(x)}{g'(x)} \tag{24}$$

provided the limit on the right exists as a finite limit or as either $+\infty$ or $-\infty$. The general conditions are that f and g are differentiable as $x \to a$ and neither $g(x)$ nor $g'(x)$ is 0 as $x \to a$.

Taylor's formula. This formula, Eq. (25), is a finite sum

$$f(x) = f(a) + \frac{f'(a)}{1!}(x - a) + \cdots + \frac{f^{(n)}(a)}{n!}(x - a)^n + R_n \tag{25}$$

expression for $f(x)$, where R_n is called the remainder. The two most useful formulas for R_n are Eq. (26), integral form, and Eq. (27), Lagrange's form. In the

$$R_n = \frac{1}{n!} \int_a^x f^{(n+1)}(t)(x - t)^n \, dt \tag{26}$$

$$R_n = \frac{f^{(n+1)}(X)}{(n+1)!}(x - a)^{n+1} \tag{27}$$

latter form, X is some number between x and a. If, for fixed a and x, $R_n \to 0$ as $n \to \infty$ $f(x)$ is said to be represented by Taylor's infinite series expansion. *See* INTEGRATION; OPERATOR THEORY; SERIES.

Angus E. Taylor

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Diffraction

The bending of light, or other waves, into the region of the geometrical shadow of an obstacle. More exactly, diffraction refers to any redistribution in space of the intensity of waves that results from the presence of an object that causes variations of either the amplitude or phase of the waves. Most diffraction gratings cause a periodic modulation of the phase across the wavefront rather than a modulation of the amplitude. Although diffraction is an effect exhibited by all types of wave motion, this article will deal only with electromagnetic waves, especially those of visible light. Some important differences that occur with microwaves will also be mentioned. For discussion of the phenomenon as encountered in other types of waves *see* ELECTRON DIFFRACTION; NEUTRON DIFFRACTION; SOUND. *See also* ELECTROMAGNETIC WAVE.

Diffraction is a phenomenon of all electromagnetic radiation, including radio waves; microwaves; infrared, visible, and ultraviolet light; and x-rays. The effects for light are important in connection with the resolving power of optical instruments. *See* RADIOWAVE PROPAGATION; X-RAY DIFFRACTION.

There are two main classes of diffraction, which are known as Fraunhofer diffraction and Fresnel diffraction. The former concerns beams of parallel light, and is distinguished by the simplicity of the mathematical treatment required and also by its practical importance. The latter class includes the effects in divergent light, and is the simplest to observe experimentally. A complete explanation of Fresnel diffraction has challenged the most able physicists, although a satisfactory approximate account of its main features was given by A. Fresnel in 1814. At that time, it played an important part in establishing the wave theory of light.

To illustrate the difference between methods of observation of the two types of diffraction, **Fig. 1** shows the experimental arrangements required to observe them for a circular hole in a screen s. The light originates at a very small source O, which can conveniently be a pinhole illuminated by sunlight. In

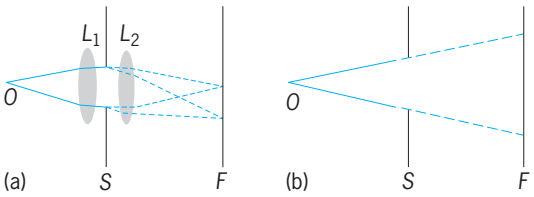

Fig. 1. Observation of the two principal types of diffraction, in the case of a circular aperture. (*a*) Fraunhofer and (*b*) Fresnel diffraction.

Fraunhofer diffraction, the source lies at the principal focus of a lens L_1, which renders the light parallel as it falls on the aperture. A second lens L_2 focuses parallel diffracted beams on the observing screen F, situated in the principal focal plane of L_2. In Fresnel diffraction, no lenses intervene. The diffraction effects occur chiefly near the borders of the geometrical shadow, indicated by the broken lines. An alternative way of distinguishing the two classes, therefore, is to say that Fraunhofer diffraction concerns the effects near the focal point of a lens or mirror, while Fresnel diffraction concerns those effects near the edges of shadow. Photographs of some diffraction patterns of each class are shown in **Fig. 2**. All of these may be demonstrated especially well by using the light beam from a neon-helium laser. *See* LASER.

Fraunhofer Diffraction

This class of diffraction is characterized by a linear variation of the phases of the Huygens secondary waves with distance across the wavefront, as they arrive at a given point on the observing screen. At the instant that the incident plane wave occupies the plane of the diffracting screen, it may be regarded as sending out, from each element of its surface, a multitude of secondary waves, the joint effect of which is to be evaluated in the focal plane of the lens L_2. The analysis of these secondary waves involves taking account of both their amplitudes and their phases. The simplest way to do this is to use a graphical method, the method of the so-called vibration curve, which can readily be extended to cases of Fresnel diffraction. *See* HUYGENS' PRINCIPLE.

Vibration curve. The basis of the graphical method is the representation of the amplitude and phase of a wave arriving at any point by a vector, the length of which gives the magnitude of the amplitude, and the slope of which gives the value of the phase. In **Fig. 3** are shown two vectors of amplitudes a_1 and a_2, pertaining to two waves having a phase difference δ of 60°. That is, the waves differ in phase by one-sixth of a complete vibration. The resultant amplitude A and phase θ (relative to the phase of the first wave) are then found from the vector sum of a_1 and a_2, as indicated. A mathematical proof shows that this proposition is rigorously correct and that it may be extended to cover the addition of any number of waves.

The vibration curve results from the addition of a large (really infinite) number of infinitesimal vectors, each representing the contribution of the Huygens secondary waves from an element of surface of the wavefront. If these elements are assumed to be of equal area, the magnitudes of the amplitudes to be added will all be equal. They will, however, generally differ in phase, so that if the elements were small but finite each would be drawn at a small angle with the

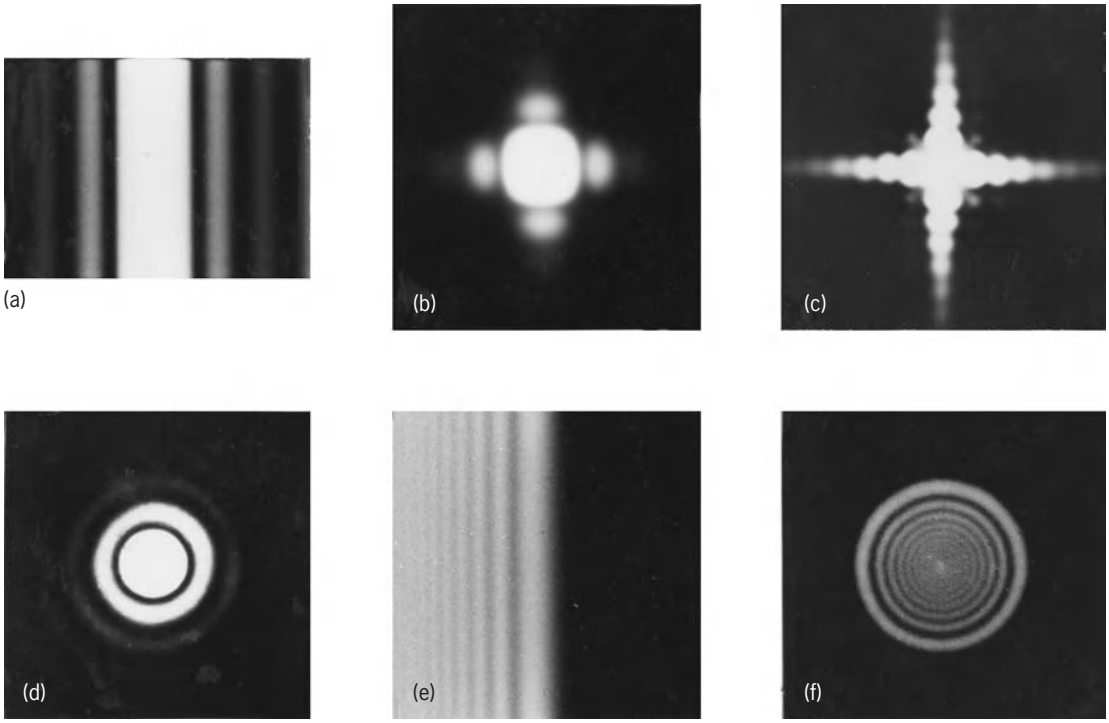

Fig. 2. Diffraction patterns, photographed with visible light. (*a*) Fraunhofer pattern, for a slit; (*b*) Fraunhofer pattern, square aperture, short exposure; (*c*) Fraunhofer pattern, square aperture, long exposure (*courtesy of F. S. Harris*); (*d*) Fraunhofer pattern, circular aperture (*courtesy of R. W. Ditchburn*); (*e*) Fresnel pattern, straight edge; (*f*) Fresnel pattern, circular aperture.

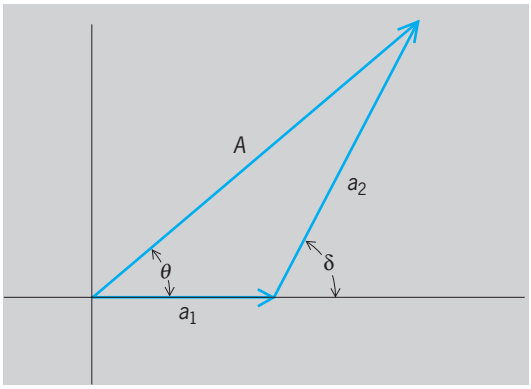

Fig. 3. Graphical addition of two amplitudes.

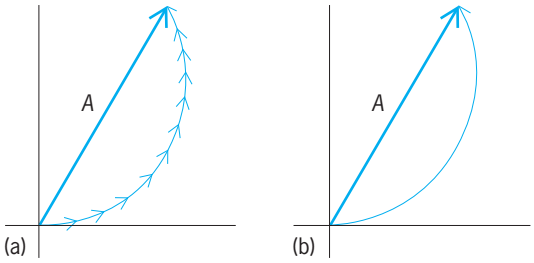

Fig. 4. Vibration curves. (*a*) Addition of many equal amplitudes differing in phase by equal amounts. (*b*) Equivalent curve, when amplitudes and phase differences become infinitesimal.

preceding one, as shown in **Fig. 4*a***. The resultant of all elements would be the vector A. When the individual vectors represent the contributions from infinitesimal surface elements (as they must for the Huygens wavelets), the diagram becomes a smooth curve, the vibration curve, shown in Fig. 4*b*. The intensity on the screen is then proportional to the square of this resultant amplitude. In this way, the distribution of the intensity of light in any Fraunhofer diffraction pattern may be determined.

The vibration curve for Fraunhofer diffraction by screens having slits with parallel, straight edges is a circle. Consider, for example, the case of a slit of width b illustrated in **Fig. 5**. The edges of the slit

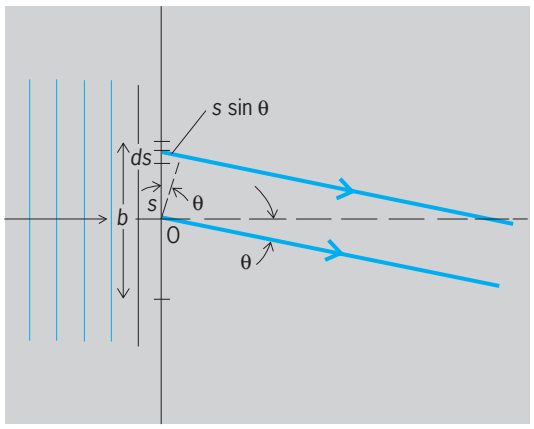

Fig. 5. Analysis of Fraunhofer diffraction by a slit.

extend perpendicular to the plane of the figure, and the slit is illuminated by plane waves of light coming from the left. If s is the distance from O of a surface element ds (ds actually being a strip extending perpendicular to the figure), the extra distance wavelet from ds must travel in reaching a point on the screen lying at the angle θ from the center is $s \sin \theta$. Since this extra distance determines the phase difference, the latter varies linearly with s. This condition necessitates that the vibration curve be a circle.

The intensity distribution for Fraunhofer diffraction by a slit as a function of the angle θ may be simply calculated as follows. The extra distances traveled by the wavelets from the upper and lower edges of the slit, as compared with those from the center, are $+(b/2) \sin \theta$ and $-(b/2) \sin \theta$. The corresponding phase differences are $2\pi/\lambda$ times these quantities, λ being the wavelength of the light. Using the symbol β for $(\pi b \sin \theta)/\lambda$, it is seen that the end points of the effective part of the vibration curve must differ in slope by $\pm\beta$ from the slope at its center, where it is taken as zero. **Figure 6** shows the form of the

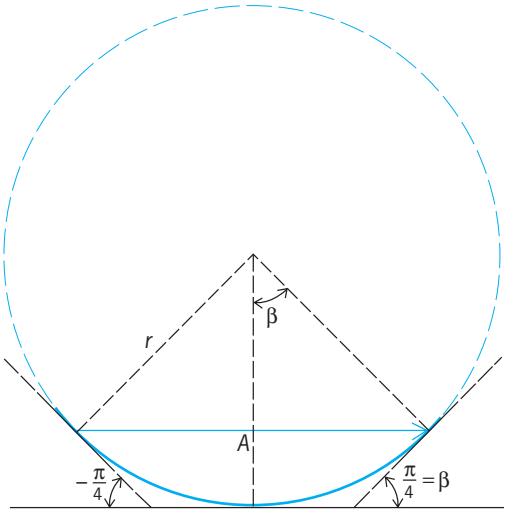

Fig. 6. Vibration curve and resultant amplitude for a particular point in Fraunhofer diffraction by a slit.

vibration curve for $\beta = \pi/4$, that is for $\sin \theta = {}^1/_4 \lambda b$. The resultant is $A = 2r \sin \beta$, where r is the radius of the arc. The amplitude A_0 that would be obtained if all the secondary waves were in phase at the center of the diffraction pattern, where $\theta = 0$, is the length of the arc. Thus, Eq. (1) holds. The intensity at any

$$\frac{A}{A_0} = \frac{\text{chord}}{\text{arc}} = \frac{2r \sin \beta}{2r\beta} = \frac{\sin \beta}{\beta} \qquad (1)$$

angle is given by Eq. (2), where I_0 is the intensity at the center of the pattern. **Figure 7** shows a graph

$$I = I_0 \frac{\sin^2 \beta}{\beta^2} \qquad (2)$$

of this function. The central maximum is twice as wide as the subsidiary ones, and is about 21 times as intense as the strongest of these. A photograph of this pattern is shown in Fig. 2*a*.

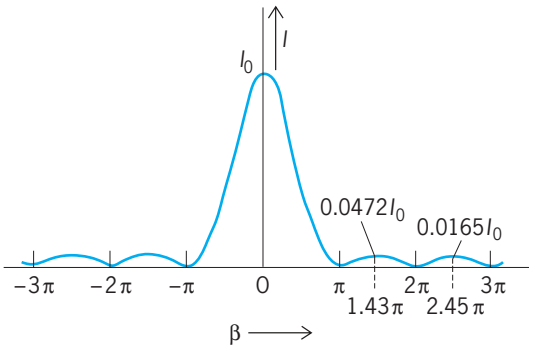

Fig. 7. Intensity distribution curve for Fraunhofer diffraction by a slit.

The dimensions of the pattern are important, since they determine the angular spread of the light behind the slit. The first zeros occur at values $\beta = (\pi b \sin \theta)/\lambda = \pm\pi$. In most cases the angle θ is extremely small, so that Eq. (3) holds. For a slit 1 mm

$$\sin\theta \approx \theta = \pm\frac{\lambda}{b} \tag{3}$$

wide, for example, and green light of wavelength 5×10^{-5} cm, Eq. (3) gives the angle as only 0.0005 radian, or 1.72 minutes of arc. The slit would have to be much narrower than this, or the wavelength much longer, for the approximation to cease to be valid.

The main features of Fraunhofer diffraction patterns of other shapes can be understood with the aid of the vibration curve. Thus for a rectangular or square aperture, the wavefront may be subdivided into elements parallel to either of two adjacent sides, giving an intensity distribution which follows the curve of Fig. 7 in the directions parallel to the two sides. Photographs of such patterns appear in Fig. 2*b* and *c*. In Fig. 2*c* it will be seen that there are also faint subsidiary maxima lying off the two principal directions. These have intensities proportional to the products of the intensities of the side maxima in the slit pattern. The fact that these subsidiary intensities are extremely low compared with that of the central maximum has an important application to the apodization of lenses, to be discussed later.

Diffraction grating. An idealized diffraction grating consists of a large number of similar slits, equally spaced. Equal segments of the vibration curve are therefore effective, as shown in **Fig. 8***a*. The resultants *a* of each segment are then to be added to give *A*, the amplitude due to the whole grating, as shown in Fig. 8*b*. The phase difference between successive elements is here assumed to be very small. As it is increased, by going to a larger angle θ, the resultant *A* first goes to zero at an angle corresponding to λ/W, where *W* is the total width of the grating. After going through numerous low-intensity maxima, *A* again rises to a high value when the phase difference between the successive vectors for the individual slits approaches a whole vibration. These small vectors *a* are then all lined up again, as they were at the center of the pattern ($\theta = 0$). The resulting strong maximum represents the "first-order spectrum," since its position depends on the wavelength. A similar condition occurs when the phase difference becomes two, three, or more whole vibrations, giving the higher-order spectra. By means of this diagram, it is possible not only to predict the intensities of successive orders for an ideal grating, but also to find the sharpness of the maxima which represent the spectrum lines. *See* DIFFRACTION GRATING.

Determination of resolving power. Fraunhofer diffraction by a circular aperture determines the resolving power of instruments such as telescopes, cameras, and microscopes, in which the width of the light beam is usually limited by the rim of one of the lenses. The method of the vibration curve may be extended to find the angular width of the central diffraction maximum for this case. **Figure 9** compares the treatments of square and circular apertures by showing, above, the elements of equal phase difference into which the wavefront may be divided, and, below, the corresponding vibration curves. For the square aperture shown in Fig. 9*a*, the areas of the surface elements are equal, and the curve forms a complete circle at the first zero of intensity. In Fig. 9*b* these areas, and hence the lengths of the successive vectors, are not equal, but increase as the center of the curve is approached,

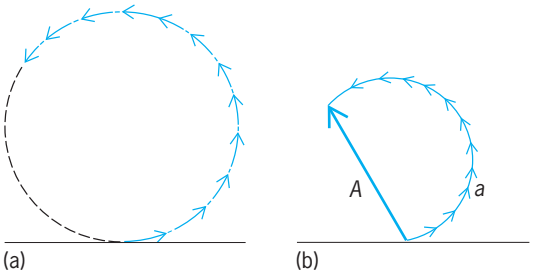

Fig. 8. Graphical analysis of diffraction grating. (*a*) Vibration curve for grating of 12 slits. (*b*) Resultant amplitude *A* formed by adding amplitudes *a* from individual slits. Each *a* represents the chord of one of the short arcs in part *a*.

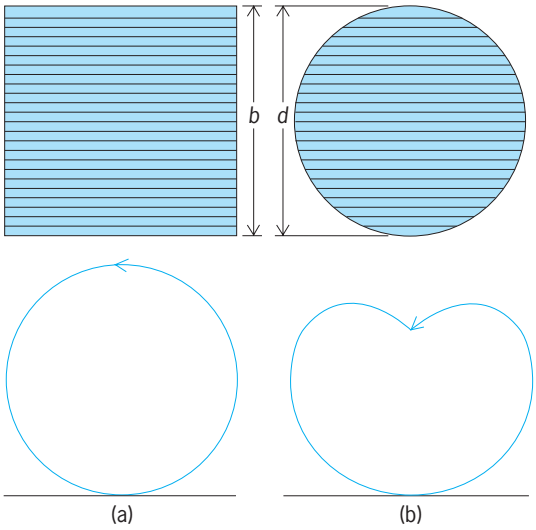

Fig. 9. Condition of vibration curves at first minimum for (*a*) a square aperture and (*b*) a circular aperture.

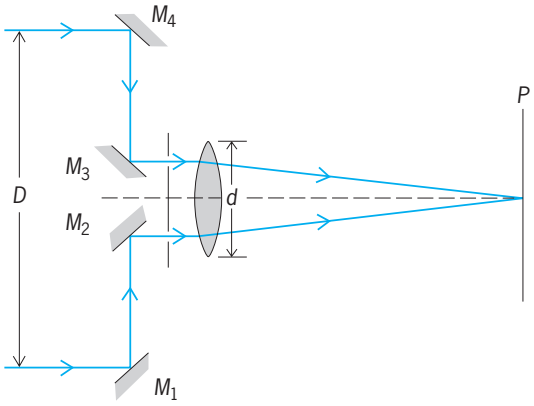

Fig. 10. Michelson stellar interferometer, which is used for measuring the diameters of stars.

and then decrease again. The result is that the curve must show somewhat greater phase differences at its extremes in order to form a closed figure. An exact construction of the curve or, better, a mathematical calculation shows that the extreme phase differences required are $\pm 1.220\pi$, yielding Eq. (4)

$$\sin\theta \approx \theta = \pm\frac{1.220\lambda}{d} \tag{4}$$

for the angle θ at the first zero of intensity. Here d is the diameter of the circular aperture. When this result is compared with that of Eq. (3), it is apparent that, relative to the pattern for a rectangle of side $b = d$, this pattern is spread out by 22%. Obviously it now has circular symmetry and consists of a diffuse central disk, called the Airy disk, surrounded by faint rings (Fig. 2*d*). The angular radius of the disk, given by Eq. (4), may be extremely small for an actual optical instrument, but it sets the ultimate limit to the sharpness of the image, that is, to the resolving power. *See* RESOLVING POWER (OPTICS).

Other applications. Among the applications of Fraunhofer diffraction, other than the calculation of resolving power, are its use in (1) the theory of certain types of interferometers in which the interfering light beams are brought together by diffraction, (2) the theory of microscopic imaging, and (3) the design of apodizing screens for lenses.

Michelson's stellar interferometer. This instrument was devised by A. A. Michelson to overcome the limitation expressed by Eq. (4). In front of a telescope are placed two fixed mirrors, shown at M_2 and M_3 in **Fig. 10**, and two others, M_1 and M_4, movable so as to vary their separation D. The light of a star is directed by these mirrors into a telescope as shown, and the diffraction patterns, of size corresponding to the aperture of the mirrors, are superimposed in the focal plane P of the lens. When D is small the resulting pattern is crossed by interference fringes. If the light source has a finite width, increasing D causes these fringes to become indistinct and eventually to disappear. For a circular disk source, such as a star, the angle subtended by the disk must be $1.22\lambda/D$ for the disappearance of the fringes. The resolving power of the telescope has thus effectively been increased in the ratio D/d. In this way angular diameters as small as 0.02 second of arc have been measured, corresponding to $D = 7$ m or 24 ft. *See* INTERFERENCE OF WAVES; INTERFEROMETRY.

Microscopic imaging. E. Abbe's theory of the microscope evaluates the resolving power of this instrument considering not only the diffraction caused by the limited aperture of the objective, but also that caused by the object itself. If this object is illuminated by coherent light, such as a parallel beam from a point source, its Fraunhofer diffraction pattern is formed in the rear focal plane of the microscope objective. Abbe took as an object a diffraction grating, and in this case the pattern consists of a series of sharp maxima representing the various orders $m = 0, \pm 1, \pm 2, \ldots$. If the light of all these orders is collected by the objective, a perfect image of the grating can be formed where the light is reunited in the image plane. In practice, however, the objective can include only a limited number of them, as is indicated in **Fig. 11**. Here only the orders $+2$ to -2 are shown entering the objective. The higher orders, involving greater values of θ, would miss the lens.

The final image must be produced by interference in the image plane of the Huygens secondary waves coming from the various orders. To obtain a periodic variation of intensity in that plane, at least two orders must be accepted by the objective. The angle θ_1 at which the first order occurs is given by $\sin\theta_1 = \lambda/d$,

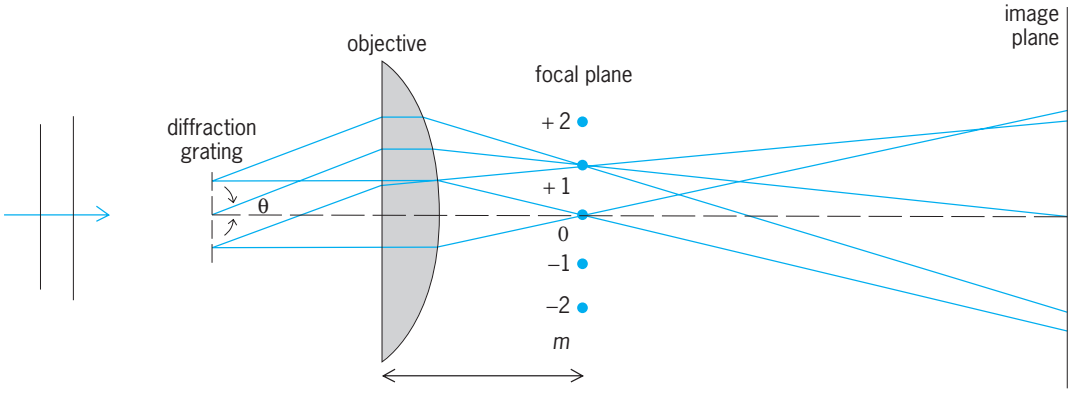

Fig. 11. Abbe's method of treating the resolving power of a microscope.

where d is the spacing of the lines in the grating. The limit of resolution of the microscope, that is, the smallest value of d that will produce an indication of separated lines in the image, may thus be found from the angular aperture of the objective. In order for it to accept only the orders 0, ±1, the lens aperture 2α must equal at least $2\theta_1$, giving $d = \lambda/\sin\alpha$. This resolving limit may be decreased by illuminating the grating from one side, so that the zero-order light falls at one edge of the lens, and that of one of the first orders at the other edge. Then $\theta_1 = 2\alpha$, and the limit of resolution is approximately $0.5\lambda/\sin\alpha$. *See* OPTICAL MICROSCOPE.

Apodization. This is the name given to a procedure by which the effect of subsidiary maxima (such as those shown in Fig. 7) may be partially suppressed. Such a suppression is desirable when one wishes to observe the image of a very faint object adjacent to a strong one. If the fainter object has, for example, only $^1/_{1000}$ of the intensity of the stronger one, the two images will have to be far enough apart so that the principal maximum for the fainter one is at least comparable in intensity to the secondary maxima for the stronger object at that point. In the pattern of a rectangular aperture it is not until the tenth secondary maximum that the intensity of the secondary maxima falls below $^1/_{1000}$ of the intensity of the principal maximum. Here it has been assumed, however, that the fainter image lies along one of the two principal directions of diffraction of the rectangular aperture, perpendicular to two of its adjacent sides. At 45° to these directions the subsidiary maxima are much fainter (Fig. 2*c*), and even the second one has an intensity of only $^1/_{3700}$, the square of the value of the second subsidiary maximum indicated in Fig. 7.

The simplest apodizing screen is a square aperture placed over a lens, the diagonal of the square being equal to the diameter of the lens. If the lens is the objective of an astronomical telescope, for example, the presence of a fainter companion in a double-star system can often be detected by turning the square aperture until the image of the companion star lies along its diagonal. Apodizing screens may be of various shapes, depending on the purpose to be achieved. It has been found that a screen of graded density, which shades the lens from complete opacity at the rim to complete transparency at a small distance inward, is effective in suppressing the circular diffraction rings surrounding the Airy disk. In all types of apodization there is some sacrifice of true resolving power, so that it would not be used if the two images to be resolved were of equal intensity.

Fresnel Diffraction

The diffraction effects obtained when the source of light or the observing screen are at a finite distance from the diffracting aperture or obstacle come under the classification of Fresnel diffraction. This type of diffraction requires for its observation only a point source, a diffracting screen of some sort, and an observing screen. The latter is often advantageously replaced by a magnifier or a low-power microscope. The observed diffraction patterns generally differ according to the radius of curvature of the wave and the distance of the point of observation behind the screen. If the diffracting screen has circular symmetry, such as that of an opaque disk or a round hole, a point source of light must be used. If it has straight, parallel edges, it is desirable from the standpoint of brightness to use an illuminated slit parallel to these edges. In the latter case, it is possible to regard the wave emanating from the slit as a cylindrical one. For the purpose of deriving the vibration curve, the appropriate way of dividing the wavefront into infinitesimal elements is to use annular rings in the first case, and strips parallel to the axis of the cylinder in the second case.

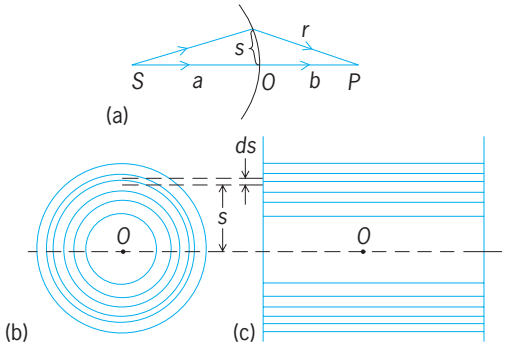

Fig. 12. Division of wavefront for constructing vibration curve for Fresnel diffraction. (*a*) Section of wave diverging from *S* and paths to screen of two secondary wavelets. (*b*) Annular zones. (*c*) Strips.

Figure 12 illustrates the way in which the radii of the rings, or the distances to the edges of the strips, must be chosen in order that the phase difference may increase by an equal amount from one element to the next. Figure 12*a* shows a section of the wavefront diverging from the source S, and the paths to the screen of two secondary wavelets. The shortest possible path is b, while r is that for another wavelet originating at a distance s above the "pole" O. Since all points on the wavefront are at the same distance a from S, the path difference between the two routes from S to P is $r - b$. When this is evaluated, to terms of the first order in s/a and s/b, the phase difference is given by Eq. (5). The phase difference across an elementary zone of radius s and width ds then becomes $d\delta = 2Csds$, so that for equal increments of δ the increment of s must be proportional to $1/s$. The annular zones and the strips drawn in this way on the spherical or cylindrical wave, respectively, looking toward the pole from the direction of P, are shown in Fig. 12*b* and *c*.

$$\delta = \frac{2\pi}{\lambda}(r - b) \approx \frac{\pi(a + b)}{ab\lambda}s^2 = Cs^2 \qquad (5)$$

For the annular zonal elements, the areas $2\pi sds$ are all equal, and hence the amplitude elements of the vibration curve should have the same magnitude. Actually, they must be regarded as falling off slowly, due to the influence of the "obliquity factor" of Huygens' principle. The resulting vibration curve is nearly, but not quite, circular, and is illustrated in **Fig. 13***a*. It

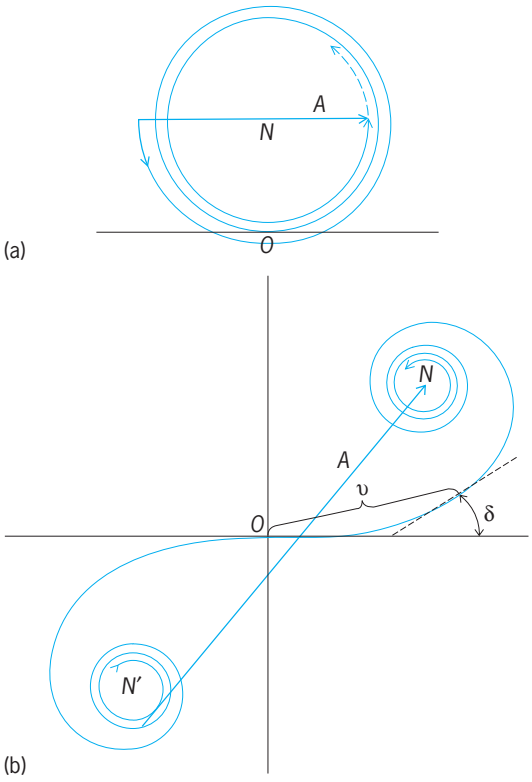

Fig. 13. Vibration curves for Fresnel diffraction patterns. (*a*) Circular division of wavefront. (*b*) Strip division (Cornu's spiral).

spirals in toward the center C, at a rate that has been considerably exaggerated in the figure. The intensity at any point P on the axis of a circular screen centered on O can now be determined as the square of the resultant amplitude A for the appropriate part or parts of this curve. The curve shown in Fig. 13*a* is for a circular aperture which exposes five Fresnel zones, each one represented by a half-turn of the spiral. The resultant amplitude is almost twice as great (and the intensity four times as great) as it would be if the whole wave were exposed, in which case the vector would terminate at C. The diffraction by other circular screens may be determined in this same manner.

Zone plate. This is a special screen designed to block off the light from every other half-period zone, and represents an interesting application of Fresnel diffraction. The Fresnel half-period zones are drawn, with radii proportional to the square roots of whole numbers, and alternate ones are blackened. The drawing is then photographed on a reduced scale. When light from a point source is sent through the negative, an intense point image is produced, much like that formed by a lens. The zone plate has the effect of removing alternate half-turns of the spiral, the resultants of the others all adding in the same phase. By putting $\delta = \pi$ and $a = \infty$ in Eq. (5), it is found that the "focal length" b of a zone plate is s_1^2/λ, where s_1 is the radius of the first zone.

Cornu's spiral. This is the vibration curve for a cylindrical wavefront, and is illustrated in Fig. 13*b*. The areas of the elementary zones, and hence the magnitudes of the component vectors of the vibration curve, decrease rapidly, being proportional to ds, and hence to $1/s$ (Fig. 12*c*). The definition of Cornu's spiral requires that its slope δ at any point be proportional to the square of the corresponding distance s measured up the wavefront [Fig. 12*a* and Eq. (5)]. The length of the spiral from the origin is proportional to s, but it is usually drawn in terms of the dimensionless variable υ, defined by $\upsilon = s\sqrt{2C/\pi}$ in the notation of Eq. (5). The coordinates of any point on the curve may then be found from tables of Fresnel's integrals.

As an example of the application of Cornu's spiral, consider the diffraction of an opaque straight edge, such as a razor edge, illuminated by light from a narrow slit parallel to the edge. At some point outside the edge of the shadow, say one which exposes three half-period strips beyond the pole, the resultant amplitude will be that labeled A in Fig. 13*b*. The intensity will be greater than that given by the amplitude NN', which represents the amplitude for the whole (unobstructed) wave. On going further away from the edge of the shadow, the tail of the vector A will move along the spiral inward toward N', and the intensity will pass through maxima and minima. At the edge of the geometrical shadow the amplitude is ON, and the intensity is just one-fourth of that due to the unobstructed wave. Further into the shadow the intensity approaches zero regularly, without fluctuations, as the tail of the vector moves up toward N. A photograph of the straight-edge pattern is shown in Fig. 2*e*.

Babinet's principle. This states that the diffraction patterns produced by complementary screens are identical. Two screens are said to be complementary when the opaque parts of one correspond to the transparent parts of the other, and vice versa. Babinet's principle is not very useful in dealing with Fresnel diffraction, except that it may furnish a shortcut method in obtaining the pattern for a particular screen from that of its complement. The principle has an important application for Fraunhofer diffraction, in parts of the field where there is zero intensity without any screen. Under this condition the amplitudes produced by the complementary screens must be equal and opposite, since the sum of the effects of their exposed parts gives no light. The intensities, being proportional to the squares of the amplitudes, must therefore be equal. In Fraunhofer diffraction the pattern due to a disk is the same as that due to a circular hole of the same size.

Diffraction of Microwaves

The diffraction of microwaves, which have wavelengths in the range of millimeters to centimeters, has been intensively studied since World War II because of its importance in radar work. Many of the characteristics of optical diffraction can be strikingly demonstrated by the use of microwaves. Microwave diffraction shows certain features, however, that are not in agreement with the Huygens-Fresnel theory, because the approximations made in that theory are no longer valid. Most of these

approximations, for example, that made in deriving Eq. (5), depend for their validity on the assumption that the wavelength is small compared to the dimensions of the apparatus. Furthermore, it is not legitimate to postulate that the wave has a constant amplitude across an opening, and zero intensity behind the opaque parts, except for the very minute waves of light.

As an example of the failure of classical diffraction theory when applied to microwaves, the results of the diffraction by a circular hole in a metal screen may be mentioned. The observed patterns begin to show deviations from the Fresnel theory, and even from the more rigorous Kirchhoff theory, when the point of observation is within a few wavelengths' distance from the plane of the aperture. Even in this plane itself there are detection effects that could not have been predicted from the earlier theories, which treat light as a scalar, rather than a vector, wave motion. An exact vector theory of diffraction, developed by A. Sommerfeld, has been applied only in a few simple cases, but the measurements at microwave frequencies agree with it wherever it has been tested. *See* LIGHT; MICROWAVE OPTICS.

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Diffraction grating

An optical device that exhibits a spatial periodic variation in one or more of the following properties: transmittance, surface profile, or refractive index. Typically, though not necessarily, the variation is one dimensional (only in one direction in the plane of the grating), and the period varies from about one-half to about one hundred times the wavelength of interest. When the variation is two dimensional, the component is normally called a diffractive optical element (DOE).

A periodic arrangement of parallel narrow slits is the simplest example of a grating. Some basic characteristics of all grating types can be understood with reference to this simple form. More commonly, gratings are periodic corrugated (grooved) surface structures, and can be used in transmission or reflection, in the latter case by coating the surface with a suitable reflective coating. The same basic principles apply in all cases.

Diffraction by periodic structures. If a parallel beam of monochromatic light of wavelength λ is incident on a periodic slit assembly of period d and slit width s (**Fig. 1***a*), then diffraction theory shows that the beam will be separated into several beams traveling in different directions. Far from the grating, the diffracted irradiance is of the form given by Eq. (1),

$$I = I_0 \left(\frac{\sin \beta}{\beta} \right)^2 \left[\frac{\sin(N\gamma)}{N \sin \gamma} \right]^2 \tag{1}$$

where I_0 is the incident irradiance, N is the number of slits, $\beta = \pi Ls/\lambda$, $\gamma = \pi Ld/\lambda$, and L is the cosine of the angle that the beam direction forms with the x axis. This function is plotted in Fig. 1*b*. The sharp main peaks, occurring along the directions indicated in Fig. 1(*a*), are called diffraction orders. *See* DIFFRACTION.

For the case of parallel slits, the distribution of power between orders depends on the ratio s/d; for the general case, it depends on the groove characteristics. In all cases, the principal peaks become sharper and the secondary oscillations weaker as the number of grooves N increases. The sharpness of the peaks can be understood if the grating is thought of as a multiple-beam interference device, where each slit or groove contributes one interfering beam. The interference picture is also helpful in understanding the position of the peaks: they will be formed in the direction where constructive interference occurs. This direction is given by the grating equation (2), where a is the angle of incidence, b is

$$\sin a \pm \sin b = m\lambda/d \tag{2}$$

the angle of diffraction, d is the grating period, and m is an integer representing the order of diffraction. If a Cartesian sign convention for angles is adopted (positive if counterclockwise from the axis of the incident beam), the negative sign applies to transmission gratings and the positive sign to reflection gratings. *See* INTERFERENCE OF WAVES.

Blazing. For a given angle of incidence, the angle of diffraction increases with $1/d$. Since the absolute value of $m\lambda/d$ is less than 2 [as follows from Eq. (2) and the fact that the absolute values of sin a and sin b are both less than 1], as d approaches the wavelength the grating equation is satisfied for fewer orders and eventually only one order. In such a case, the diffracted energy can be concentrated mostly in a single order, which is generally desirable. With coarser gratings, many orders are allowed. However, even in this case an appropriate groove profile can also concentrate the light in a desired order. A grating with such a profile is called blazed. The principle of blazing is explained in **Fig. 2**. Light incident on a facet tends to be refracted or reflected according to Snell's law or the law of reflection, respectively. If the direction of refraction or reflection coincides with an allowed direction of diffraction, the energy is concentrated in that direction. The facet angle is called the blaze angle θ; for reflection gratings, it is related to the angles of incidence and diffraction through $\theta = (a + b)/2$. Since the diffraction angle changes with wavelength, only one wavelength (the blaze wavelength) will diffract at the desired angle once the angle of incidence is fixed. *See* REFLECTION OF ELECTROMAGNETIC RADIATION; REFRACTION OF WAVES.

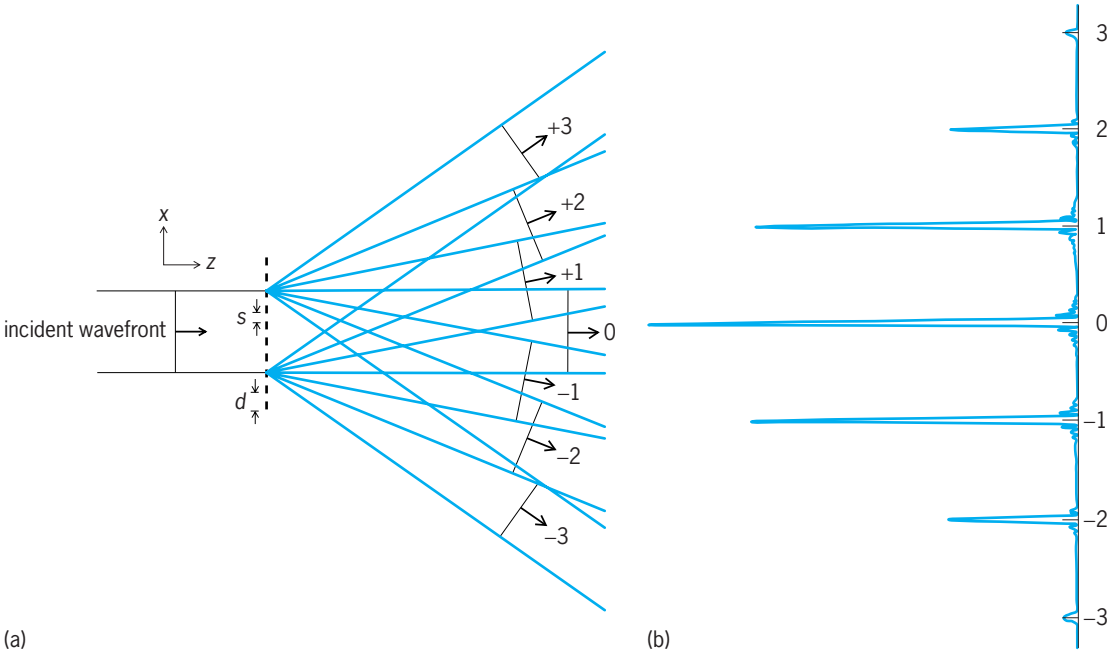

Fig. 1. Behavior of a light beam incident on a periodic slit pattern (grating). (*a*) Splitting of the beam into several beams traveling along different directions. (*b*) Far from the grating, these beams correspond to sharp peaks in the light distribution, called orders of diffraction. The distribution shown here is obtained for $N = 20$ slits and a ratio $d/s = 3.5$.

Dispersion and related properties. Though diffraction gratings can be used with monochromatic radiation, by far their most common use is to place them in the path of a light beam with the aim of dispersing the light, that is, separating it into its component wavelengths. The dispersive properties of a grating are characterized by its angular dispersion D, which can be expressed by Eq. (3), where the negative signs

$$D = \frac{db}{d\lambda} = \frac{\sin b \pm \sin a}{\lambda \cos b} = \pm \frac{m}{d \cos b} \qquad (3)$$

apply to transmission gratings. However, only the absolute value of D is significant. This equation shows that the same dispersion can be achieved with either a coarse grating used in high order or a fine grating used in low order.

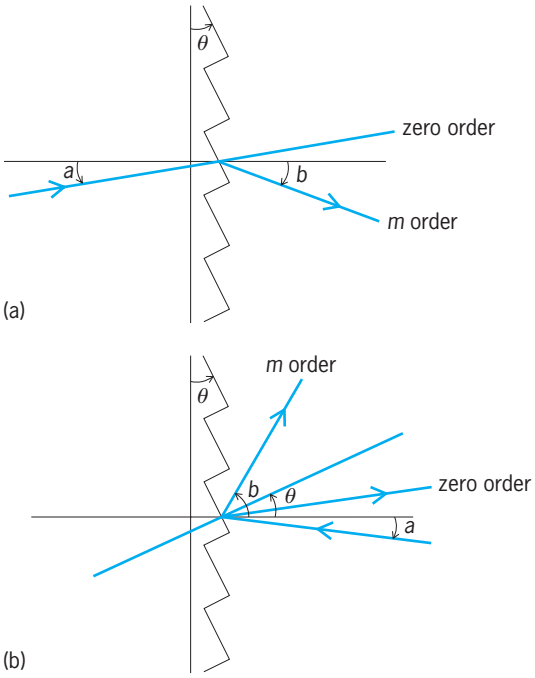

Fig. 2. Geometry of diffraction from a (*a*) transmission grating, (*b*) reflection grating.

The free spectral range (FSR) of a grating is the portion of the spectrum over which there is no overlap of wavelengths from adjacent orders, and is given as λ/m, where λ is here the lowest wavelength in the range of interest. Larger free spectral range is therefore achieved with a lower order m. This reduces the need for order-sorting filters or cross-dispersing elements to separate the orders.

The resolving power R of a grating is defined as the mean wavelength λ divided by the minimum resolvable wavelength difference $\Delta\lambda$; $R \equiv \lambda/\Delta\lambda$. It is given by $R = mN$, where N is the total number of illuminated grooves. This is independent of groove period, but insofar as there is an overall beam size limitation, a finer-pitch grating will provide more grooves and hence higher resolution.

Diffraction efficiency. Diffraction efficiency is defined as the percentage of the incident energy that is concentrated in the desired order. It can reach values above 90%, or even above 99% in special cases, but it is typically strongly wavelength-dependent. The interaction of light with coarse, shallow gratings can be described through scalar diffraction theory. Such gratings therefore show very little difference in efficiency between p (parallel to the grooves) and s (perpendicular to the grooves) polarizations. In this

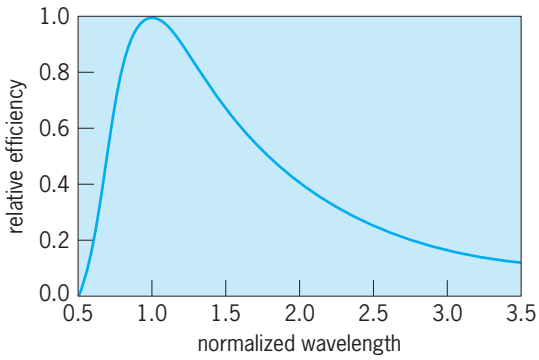

Fig. 3. Relative spectral efficiency of a coarse blazed grating as a function of normalized wavelength λ/λ_0.

case, the relative spectral efficiency takes the functional form given by Eq. (4), where λ_0 the blaze wavelength. This function is plotted in **Fig. 3** for $m = 1$. For finer gratings, the *p*-polarization relative efficiency generally retains a similar shape, while the *s*-polarization shows sharp peaks and troughs called Wood's anomalies. The efficiency curve shape can be significantly different for nonblazed gratings and can also be tailored to some extent through modification of the groove shape. *See* POLARIZED LIGHT.

$$I/I_0 = \left\{ \frac{\sin\left[\pi\left(\frac{\lambda_0}{\lambda} - m\right)\right]}{\pi\left(\frac{\lambda_0}{\lambda} - m\right)} \right\}^2 \qquad (4)$$

Parasitic light. All light that is not absorbed or diffracted in the direction of a diffraction order is parasitic. There are two main types of parasitic light: ghosts and satellites, and scatter. Ghosts and satellites are concentrations of energy in directions other than the main diffraction orders; the distinction between the two is that satellites are formed very close to a main order. Both types arise from periodic errors in grating spacing or any other periodicity introduced by the fabrication method. Diffuse or homogeneous scatter is a continuous background light, spread more or less uniformly across the angle. Its main source is the microroughness of the groove surface. In-plane scatter is light scattered in the plane of dispersion, caused mostly by random variations in groove period and depth. In several spectroscopic applications, the absence or reduction of parasitic light is of paramount importance and dictates the choice of grating.

Types of gratings. Gratings can be classified according to the mode of use as transmission or reflection. They can also be classified as surface or volume gratings, depending on where diffraction takes place. In volume gratings, the effective grooves are formed by variation of the refractive index inside the material. Volume gratings are further classified as thin or thick, with a thick (also called Bragg) grating being one that has a thickness greater than about 10 grating periods. Although diffraction is the operating principle of all these types, the term "diffraction gratings" most often refers to surface gratings used for dispersing the light. The efficiency of volume Bragg gratings can approach 100% and the width of the spectral efficiency curve can be quite narrow. These characteristics make them desirable as filters or highly efficient, wavelength-selective mirrors. Gratings are also classified according to the fabrication technique. *See* INTERFERENCE FILTER.

Fabrication techniques. Starting in the twentieth century, gratings were first fabricated by ruling, which involves repetitively running a diamond cutting tool along the surface of a metal or glass substrate. Extremely high precision is required in the motion of the tool to avoid artifacts, such as caused by sideways motion or periodic errors in the translation mechanism. An original ruled grating, called a master, is often replicated by the use of a resin that conforms to the master grating surface before hardening. Replicas can be superior to the master at least in terms of possessing smoother grooves, leading to a reduction in scattered light. The groove shape of ruled gratings is triangular (blazed). When gratings must be formed on a curved substrate, it is very difficult to maintain a constant blaze angle, as the cutting tool angle varies relative to the substrate. This results in significant variation of diffraction efficiency across the face of the grating. Partial mitigation of this effect is achieved through multipanel gratings, with different tool angles between panels. It is difficult, however, to maintain phase coherence between the panels, and the result is typically a grating that has the resolving power corresponding to the number of rulings in only one panel.

The development of lasers made possible the fabrication of holographic (or, more properly, interference) gratings, which are made by recording the interference fringe pattern from two coherent light beams onto a suitable recording medium such as a photoresist, followed by development as necessary to form a surface pattern. The groove shape of holographic gratings is typically sinusoidal; however, development techniques can change the groove depth to advantage, and ion-milling techniques can change the groove shape. At high groove densities, holographic gratings are advantageous in terms of their potential for producing minimal scatter, as well as the flexibility to record on an arbitrary (nonplanar) surface shape. Variation of efficiency across the grating face is not observed; however, the grooves are not blazed. In addition, holographic gratings can be formed with nonstraight grooves if the interfering beams are not planar. In that case they can be used to focus a beam as well as to correct the aberrations of the optical system. So-called aberration-correcting gratings which present a slow variation in a second direction blur the distinction between diffraction gratings and diffractive optical elements, but are best classified as gratings. *See* ABERRATION (OPTICS); HOLOGRAPHY; LASER.

The development of directed-energy beam tools in addition to lasers has resulted in new fabrication techniques that provide greater flexibility by controlling the groove depth, shape, and period not only to conform to desired patterns but also to vary along the extent of the grating. Electron-beam lithography is the most advanced of these alternative techniques

at present. Like holographic gratings, these gratings can also be recorded on a variety of nonplanar substrates as well as have nonstraight grooves, while providing greater flexibility in patterning the groove shape. The inherent accuracy of lithographic tools permits the fabrication of fully coherent multipanel gratings on curved substrates. However, these are typically limited in size. Where large-area gratings are needed, holographic techniques are preferred. *See* MICROLITHOGRAPHY.

Use of gratings. A monochromator is a device that produces monochromatic (or narrow-band) radiation of any desired mean wavelength within a certain range. This is accomplished by dispersing the light of a broadband source and placing a slit or narrow aperture to select a small portion of the spectrum. By rotating the grating (changing the angle of incidence), the wavelength is swept across the stationary slit. A typical grating mounting for a monochromator is the Czerny-Turner arrangement, comprising a collimating and focusing mirror (**Fig. 4**).

A spectrometer is a device that records the spectrum of an object by dispersing the light from the object and placing a photodetector array at the output. Thus the spectrometer uses no output slit but must retain the input slit or pinhole to maintain a unique angle of incidence on the grating. When the image of a monochromatic point source is sufficiently free from aberrations, it is possible to obtain simultaneously many independent spectra of points along the input slit by using a two-dimensional array detector at the output. The detector then records the spectral information in the perpendicular direction to the slit and spatial information in the parallel direction to the slit. Such an arrangement is called an imaging spectrograph or spectrometer.

Several alternative spectrometer grating mounts are possible, in addition to the Czerny-Turner one. Concave grating mounts make use of the imaging property of a grating formed on a concave spherical surface of radius R. The entrance slit is placed on a circle of radius $R/2$ tangent to the grating at the apex, in which case the image is found on the same circle (the Rowland circle). However, the aberrations of this arrangement are severe, and modifications of the classical form using aberration-correcting holographic gratings are advantageous. For maximum resolution the Littrow mounting is needed, in which

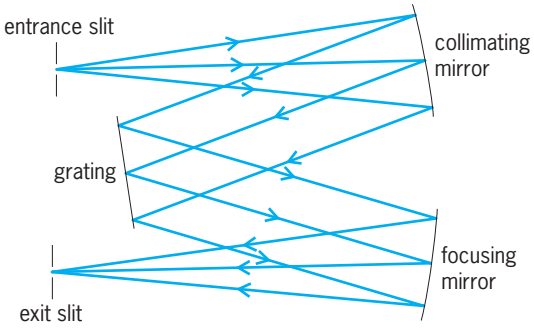

Fig. 4. Typical Czerny-Turner grating monochromator mount. The grating rotates about an axis perpendicular to the paper.

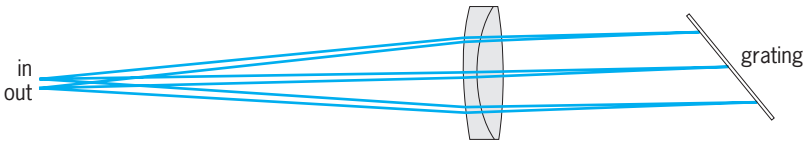

Fig. 5. Example of a grating mounted in Littrow configuration, where the angles of incidence and diffraction are approximately equal. Input and output beams are shown slightly separated in the plane of the paper, though they may also be placed above and below that plane.

the angles of incidence and diffraction are approximately equal (**Fig. 5**). A coarse grating used in the Littrow mode and in a high order of diffraction is called an echelle. *See* ASTRONOMICAL SPECTROSCOPY.

The monochromatic image of a straight slit is inherently curved in flat grating and Rowland circle–type instruments and therefore is not a good match for array detectors. Such curvature, as well as other aberrations, is well controlled in the Offner spectrograph, which uses a convex grating and concentric optics to achieve a high degree of aberration correction with low slit distortion. This makes possible the use of long slits. *See* SPECTROGRAPH; SPECTROSCOPY.

Modern developments and applications. Fabrication techniques continue to improve and new techniques are still appearing, most notably ion etching and lithography. In addition, innovative uses of gratings and new applications are constantly emerging. Imaging spectroscopy is becoming a mainstay technique in Earth and planetary remote sensing. Volume phase gratings are beginning to be considered for astronomical telescope spectrographs. Large-area gratings, assembled from several smaller ones, are needed for laser fusion applications, necessitating techniques for extremely precise assembly. Grating-based wavelength multiplexing in optical communications is an active area of development. Tunable laser cavities can utilize transmission or reflection gratings. *See* NUCLEAR FUSION; OPTICAL COMMUNICATIONS.

P. Mouroulis

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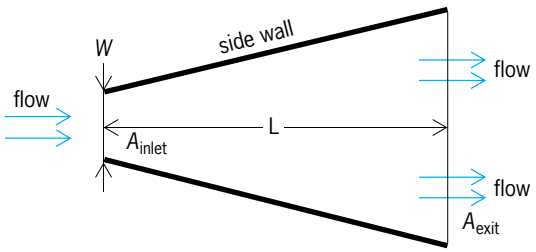
Diffuser

A device in which the kinetic energy of a flow is recovered, thus giving a static pressure rise. A simple example is a diverging passage (see **illus.**) with flow entering the small end and leaving the large end. For an incompressible flow (density, ρ, is constant) the mass conservation principle, Eq. (1), indicates

$$m = \rho V A \tag{1}$$

that the velocity (V) will decrease through the diffuser as the area (A) increases, since the mass flow rate (m) is constant. It follows from the Bernoulli



Standard diffuser geometry.

principle, Eq. (2), and from the fact that the total

$$p_0 = p + \tfrac{1}{2}\rho V^2 \tag{2}$$

pressure (p_0) changes but slightly, that the static pressure (p) rises in the diffuser. All subsonic diffusers, whether incompressible or not, experience streamtube area increase as kinetic energy is converted to static pressure rise. Supersonic diffusers operate differently since changes in density become very significant. Many different geometries are possible for diffusers. *See* BERNOULLI'S THEOREM; CONSERVATION OF MASS; SUPERSONIC DIFFUSER.

Applications. Diffusers serve many vital roles. Many fluid-dynamic processes exhibit excessive kinetic energy at some point in their operating cycle. This kinetic energy can be recovered as a useful static pressure rise by the implementation of a diffuser. For example, hydraulic turbines installed for hydroelectric plant operation invariably use a draft tube, that is, an exhaust diffuser. Likewise, ground-based gas turbines for electric power production or general-purpose expansion turbines for waste-gas expansion generally employ an exhaust diffuser for extra pressure recovery. Compressors and pumps invariably generate excessive kinetic energy as work is done by their rotors, and diffusers are required to recover the kinetic energy as a static pressure rise. The basic elements of fluid transfer include pipes, elbows, valves, and diffusers. *See* COMPRESSOR; GAS TURBINE; HYDRAULIC TURBINE; PUMP.

Performance. The performance of a diffuser is described by a static pressure recovery coefficient, C_p. In the case of incompressible flow, this coefficient is equal to the fraction of the inlet kinetic energy that is recovered or converted to a static pressure rise in the diffuser. Values of C_p cannot exceed unity (1.0) for a diffuser but can reach zero or negative numbers for diffusers whose design is so poor that the flow must be pushed through, that is, accelerated.

Diffuser performance is frequently plotted on a performance map, on which contours of recovery (curves connecting points with equal values of C_p) are shown in terms of important geometric parameters, including the area ratio (AR, the ratio of exit area to inlet area, A_{exit}/A_{inlet}) on the ordinate and the length-to-width parameter (the ratio of diffuser length to inlet width, L/W) on the abscissa. Inlet aerodynamic blockage is an important variable which establishes the effective flow-area blockage at the diffuser inlet resulting from low-momentum shear flow near the walls. Many such maps exist from which good estimates of performance can be established.

Ideal limit. It is instructive to consider the ideal limit of a diffuser. In the ideal case, the total pressure is unchanged between the inlet and the outlet. If the Bernoulli equation (2) is used to compute the static pressure rise, then Eq. (3) is obtained for the ideal recovery coefficient after the mass-conservation equation is employed. Hence, it is clear that $C_p = 1.0$ is achievable only for both ideal flow and an infinite area ratio. As examples, for $AR = 4$, $C_{p,\text{ideal}} = 0.9375$, and for $AR = 2.0$, $C_{p,\text{ideal}} = 0.75$. Of course, the ideal case can be approached only in the limit.

$$C_{p,\text{ideal}} = 1 - \frac{1}{(AR)^2} \tag{3}$$

Effectiveness. Diffuser effectiveness (η) is the ratio of the actual static pressure recovery coefficient to the ideal recovery coefficient. Levels of η in the range 0.5–0.8 are common in industrial practice if sensible guidelines are followed. Geometry is important, and diffuser divergence angles must be limited to small angles of about 7–9°. Lengths of about 12–18 times the inlet height are needed to get good diffusion without significant flow separation (reversal). Inlets with the least flow distortion possible are desired to achieve high levels of C_p and η. Inlet swirl may help or hinder performance; moderate turbulence can help diffusers of larger divergence angle. Curvature against the axis of a diffuser will invariably degrade performance. When sensible practice is employed, effectiveness levels of about 0.7–0.8 are common; higher values are possible for unusual designs, and lower values are common for poor designs as may result when space is limited and compromises are necessitated. *See* TURBULENT FLOW.

Assessment. The above guidelines permit tentative assessment of expected diffuser performance. For a given diffuser, the area ratio may be calculated first; and then the ideal static pressure recovery coefficient, using Eq. (3) for rectilinear diffusers and alternative equations for other cases, may be computed next. Effectiveness can be estimated from the guidelines given above subject to diffuser geometry and inlet fluid-dynamic conditions.

Losses. Losses in the diffuser can be related to the pressure recovery. In the case of incompressible flow, the loss coefficient can be defined as the ratio of the total pressure drop through the diffuser to the inlet kinetic energy. An approximate relationship can be developed from the principles given above whereby the loss coefficient equals the difference between the ideal static pressure recovery coefficient and the actual recovery coefficient. By using the basic concepts presented above concerning pressure recovery, including the ideal pressure recovery, a very sensible estimate of losses within the diffuser, that is, the degradation of total pressure, can be estimated. *See* FLUID FLOW; FLUID-FLOW PRINCIPLES.

David Japikse

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Diffusion

The transport of matter from one point to another by random molecular motions. It occurs in gases, liquids, and solids.

Diffusion plays a key role in processes as diverse as permeation through membranes, evaporation of liquids, dyeing textile fibers, drying timber, doping silicon wafers to make semiconductors, and transporting of thermal neutrons in nuclear power reactors. Rates of important chemical reactions are limited by how fast diffusion can bring reactants together or deliver them to reaction sites on enzymes or catalysts. The forces between molecules and molecular sizes and shapes can be studied by making diffusion measurements. *See* CELL PERMEABILITY; EVAPORATION; INTEGRATED CIRCUITS; SEMICONDUCTOR; THERMAL NEUTRONS.

Fluids

Molecules in fluids (gases and liquids) are constantly moving. Even in still air, for example, nitrogen and oxygen molecules ricochet off each other at bullet speeds. Molecular diffusion is easily demonstrated by pouring a layer of water over a layer of ink in a narrow glass tube. The boundary between the ink and water is sharp at first, but it slowly blurs as the ink diffuses upward into the clear water. Eventually, the ink spreads evenly along the tube without any help from stirring.

Diffusion equations. Adolph Fick discovered the mathematical laws of diffusion in 1855. He found that the rate of diffusion of a substance is proportional to the gradient in its concentration (C). This is expressed in his first law, Eq. (1), where J is

$$J = -D\frac{\partial C}{\partial z} \tag{1}$$

the flux density, the amount of substance diffusing parallel to the z coordinate axis per unit time per unit area. The minus sign reflects the natural tendency of substances to diffuse down concentration gradients, from higher to lower concentrations. Similar equations can be written for diffusion along the x and y coordinates in three-dimensional space.

The quantity D is called the diffusion coefficient or the diffusivity. It measures the flux produced by a given concentration gradient. If the concentration is expressed in units of moles per cubic centimeter and the flux in moles per square centimeter per second, the units of D are square centimeters per second. Mass concentrations (grams per cubic centimeter) and mass fluxes (grams per square centimeter per second) are also used. Since J is the rate of diffusion per unit area, the amount of substance diffusing per second through area A is AJ.

Fick's law for the flow of matter in a concentration gradient is analogous to Fourier's law for heat conduction in a temperature gradient. *See* CONDUCTION (HEAT).

Fick's second law. Diffusion changes the distribution of molecules as time passes. Each second, $AJ(z)$ moles of substance diffuse into a slab of solution of area A and thickness Δz (**Fig. 1**) across its left face. Simultaneously, $AJ(z + \Delta z)$ moles per second diffuse out of the slab across its right face. The net rate of diffusion into the slab is $AJ(z) - AJ(z + \Delta z)$.

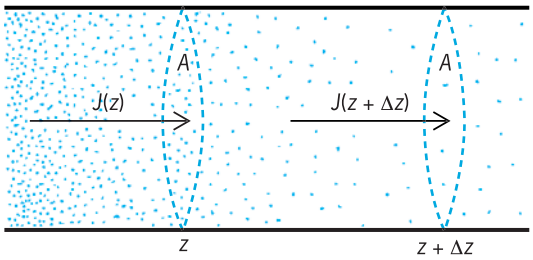

Fig. 1. Diffusion in a tube of cross-sectional area *A*. The net rate of diffusion into the fluid between *z* and *z* + Δ*z* is *AJ*(*z*) − *AJ*(*z* + Δ*z*).

Dividing the rate of diffusion into the slab by its volume $A\Delta z$ gives the rate of change of concentration, as in Eq. (2), and combining Eqs. (1) and (2) gives Fick's second law of diffusion, Eq. (3). As

$$\frac{\Delta C}{\Delta t} = -\frac{AJ(z + \Delta z) - AJ(z)}{A\Delta z} \tag{2}$$

$$\frac{\partial C}{\partial t} = -\frac{\partial J}{\partial z} = \frac{\partial}{\partial z}\left[D\frac{\partial C}{\partial z}\right] \tag{3}$$

time passes, the concentration changes in proportion to the change in the concentration gradient along the diffusion path. This illustrates diffusion's fundamental role: smoothing out gradients in composition.

Diffusion coefficients can be treated as constants if the concentration gradients are not too large. The simpler version of Fick's second law, Eq. (4), then applies

$$\frac{\partial C}{\partial t} = D\frac{\partial^2 C}{\partial z^2} \tag{4}$$

This equation is used to predict the concentrations of diffusing substances in hundreds of different applications. Slightly different equations are used for diffusion in cylindrical or spherical geometries. *See* DIFFERENTIAL EQUATION.

Onsager's equation. Just as positive charge flows from higher to lower electric potential, molecules diffuse from regions of higher to lower chemical potential (a thermodynamic property). Gradients in the chemical potential, not concentration gradients, are the true driving forces for diffusion. A more fundamental description of diffusion states that the flux of a diffusing substance is proportional to the gradient in its chemical potential, as in Eq. (5).

$$J = -L\frac{\partial \mu}{\partial z} \tag{5}$$

The thermodynamic diffusion coefficient L is proportional to the mobility of a diffusing substance, the ease with which it moves through a fluid. Small molecules in low-viscosity fluids have the largest mobilities. Fick's diffusion coefficient D equals $L\partial\mu/\partial C$.

This result shows that diffusion coefficients are not purely kinetic quantities because they depend on thermodynamic properties too. *See* CHEMICAL THERMODYNAMICS.

Diffusion processes. These comprise chemical interdiffusion, self-diffusion, and intradiffusion.

Chemical interdiffusion. By far the most frequently studied diffusion process is the molecular mixing of two different chemical substances. An example is the interdiffusion of oxygen and nitrogen. At first glance, there are two diffusion coefficients to consider, one for each substance. However, there is no bulk movement of fluid during diffusion. This restriction leads to the surprising result that the interdiffusion of two substances is described by a single diffusion coefficient, D: $J_1 = -D\partial C_1/\partial z$ and $J_2 = -D\partial C_2/\partial z$. The quantity D is called the interdiffusion coefficient or the mutual diffusion coefficient.

Gas molecules travel over distances of many molecular diameters before being deflected by collisions with other molecules. This gives gas mixtures the largest interdiffusion coefficients, approximately 0.1 $cm^2\ s^{-1}$ at room temperature and atmospheric pressure. A gas molecule diffuses about 0.1–0.5 cm in 1 s. Diffusion in liquids is about 10,000 times slower because the molecules are packed more closely. Solid-state diffusion is even slower because the molecules are locked in position by their neighbors, except for infrequent jumps.

Self-diffusion. Diffusion occurs in pure substances too, and is called self-diffusion. It can be studied by tagging some of the diffusing molecules with isotopes. For example, the self-diffusion coefficient of water is measured by releasing trace amounts of water labeled with radioactive tritium ($^3H^1HO$) at one end of a tube filled with ordinary water (^1H_2O) and then monitoring the spread of radioactivity along the tube. Magnetized molecules also serve as tagged species. *See* ISOTOPE; NUCLEAR MAGNETIC RESONANCE (NMR).

Intradiffusion. The term intradiffusion refers to the interchange of tagged and untagged species in systems of uniform chemical composition. To measure the intradiffusion coefficient of benzene in a benzene-heptane mixture containing 1 mole of benzene per liter, a 1-mole-per-liter solution of benzene in heptane is brought into contact with another 1-mole-per-liter benzene solution in which a portion of the total benzene molecules are tagged with radioactive ^{14}C.

Gases. A number of techniques are used to measure diffusion in gases. In a two-bulb experiment, two vessels of gas are connected by a narrow tube through which diffusion occurs. Diffusion is followed by measuring the subsequent changes in the composition of gas in each vessel. Excellent results are also obtained by placing a lighter gas mixture on top of a denser gas mixture in a vertical tube and then measuring the composition along the tube after a timed interval.

Rates of diffusion in gases increase with the temperature (T) approximately as $T^{3/2}$ and are inversely proportional to the pressure. The interdiffusion coefficients of gas mixtures are almost independent of the composition.

Kinetic theory shows that the self-diffusion coefficient of a pure gas is inversely proportional to both the square root of the molecular weight and the square of the molecular diameter. Interdiffusion coefficients for pairs of gases can be estimated by taking averages of the molecular weights and collision diameters. Kinetic-theory predictions are accurate to about 5% at pressures up to 10 atm (1 megapascal). Theories which take into account the forces between molecules are more accurate, especially for dense gases. *See* KINETIC THEORY OF MATTER.

Liquids. The most accurate diffusion measurements on liquids are made by layering a solution over a denser solution and then using optical methods to follow the changes in refractive index along the column of solution. Excellent results are also obtained with cells in which diffusion occurs between two solution compartments through a porous diaphragm. Many other reliable experimental techniques have been devised.

Room-temperature liquids usually have diffusion coefficients in the range $0.5\text{–}5 \times 10^{-5}\ cm^2\ s^{-1}$. Diffusion in liquids, unlike diffusion in gases, is sensitive to changes in composition but relatively insensitive to changes in pressure. Diffusion of high-viscosity, syrupy liquids and macromolecules is slower. The diffusion coefficient of aqueous serum albumin, a protein of molecular weight 60,000 atomic mass units, is only $0.06 \times 10^{-5}\ cm^2\ s^{-1}$ at 25°C (77°F).

When solute molecules diffuse through a solution, solvent molecules must be pushed out of the way. For this reason, liquid-phase interdiffusion coefficients are inversely proportional to both the viscosity of the solvent and the effective radius of the solute molecules. Accurate theories of diffusion in liquids are still under development. *See* VISCOSITY.

Applications. Diffusion is involved in a great diversity of processes.

Tracer diffusion. If tracer molecules are released in a fluid in which there is no convection, such as stirring, Fick's second equation can be integrated to predict how rapidly the tracer molecules spread out by random thermal motions.

If tracer molecules are released at position $z = 0$ in a tube of cross-sectional area A, they will cluster near the origin initially, and slowly disperse (**Fig. 2***a*). The bell-shaped concentration profiles are identical to the gaussian distribution of random errors employed in statistical sciences. This is not accidental, but a direct consequence of the random motions of diffusing molecules. In fact, the standard deviation of the tracer profile, a measure of its width, is $\sqrt{2Dt}$ after diffusion for time t. *See* STATISTICS.

Mixing without convection. Two fluids of different composition can be rapidly mixed by stirring. Without convection, the rate of mixing is diffusion controlled and much slower. As time passes, interdiffusion

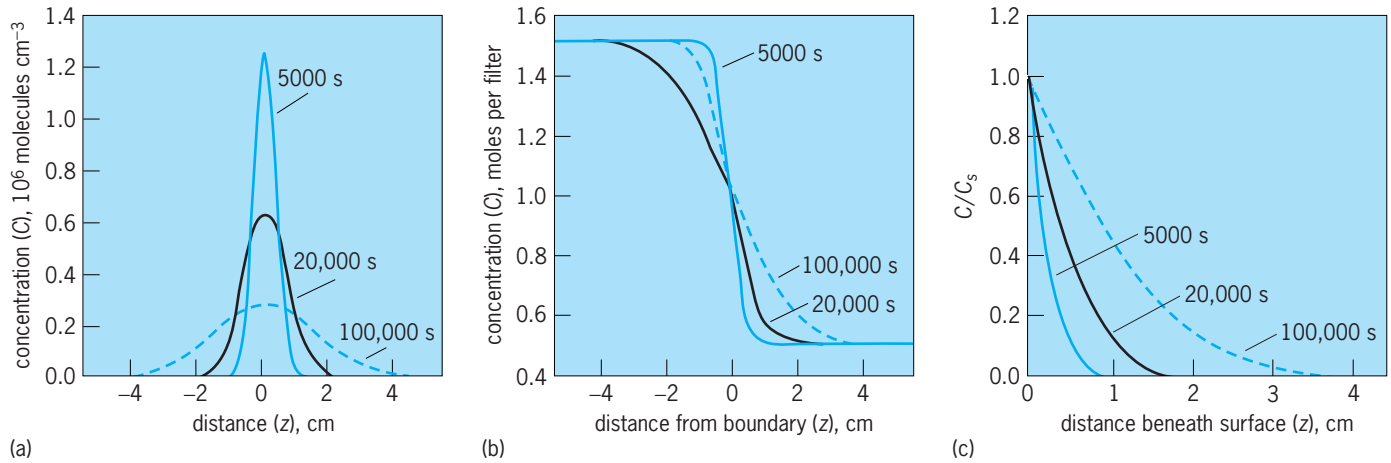

Fig. 2. Concentration profiles. (*a*) Diffusion of 10^6 tracer molecules in a tube of 1-cm^2 cross-sectional area. The molecules are initially located at position $z = 0$. $D = 1 \times 10^{-5}$ cm^2 s^{-1}. (*b*) Interdiffusion of a solution of concentration 1.5 moles per liter into a solution of concentration 0.5 mole per liter across a sharp initial boundary. (*c*) Absorption of a gas into a convection-free liquid. Concentration (C) of dissolved gas is given in terms of saturation concentration at liquid surface.

gradually smooths out a sharp initial boundary between two solutions (Fig. 2*b*).

Gas absorption. Concentration profiles can also be measured for the absorption and diffusion of a soluble gas into a convection-free liquid (Fig. 2*c*). The surface of the liquid is saturated with dissolved gas. After exposure for time t, the dissolved gas penetrates the absorbent to an average depth of about $\sqrt{Dt}$.

The diffusion flux at the surface of the liquid gives the rate of dissolution of gas per unit surface area of liquid. This rate is proportional to the solubility of the gas and the square root of the diffusion coefficient. The adsorption process slows down as time passes as $t^{-1/2}$. Identical results are obtained for the dissolution of a solid into a stagnant liquid.

Brownian motion. Colloidal particles suspended in a liquid move by a series of small, chaotic jumps. This kind of diffusion is called brownian motion. It was first observed by looking at suspensions of pollen grains in water under a microscope.

A. Einstein analyzed the motion of a randomly jumping particle. He showed that a particle making v jumps per second of length ℓ has diffusion coefficient $v\ell^2/2$. This result provides a useful physical model of the diffusion process. It also accounts for the slowness of diffusion in liquids relative to gases: The distance between collisions, ℓ, is relatively small for liquids. Diffusion in solids is even slower because both the step size and jump frequency are small.

If a particle is allowed to diffuse freely in one dimension, its root-mean-square displacement after time t is $\sqrt{2Dt}$. This result is frequently used to estimate the distances covered by diffusing molecules. A molecule with diffusion coefficient 1×10^{-5} cm^2 s^{-1} diffuses about 1 cm in a day but only about 25 cm in a year. The root-mean-square displacement for three-dimensional diffusion is $\sqrt{6Dt}$ (**Fig. 3**). *See* BROWNIAN MOVEMENT; COLLOID.

Porous materials. Diffusion through a porous solid is slower than diffusion through an equal volume of homogeneous fluid for two reasons. First, the diffusing molecules must travel greater distances to move around solid obstructions. Second, diffusion can occur only through the pore fraction of the total cross-sectional area.

A practical approach to this problem is to treat the porous medium as a homogeneous phase with an effective diffusion coefficient, D_{eff}. Values of D_{eff} are typically 3–10 times smaller than the diffusion coefficient D for the pore fluid. The D_{eff}/D ratio depends on the void fraction and details of the pore network, such as the number of interconnections and dead ends.

Effusion. At low pressures, gas molecules can escape through a pinhole in the wall of a container without colliding with other molecules. This process is called effusion. It is used to measure the vapor pressure of solids and liquids at high temperatures. The rate of effusion is inversely proportional to the

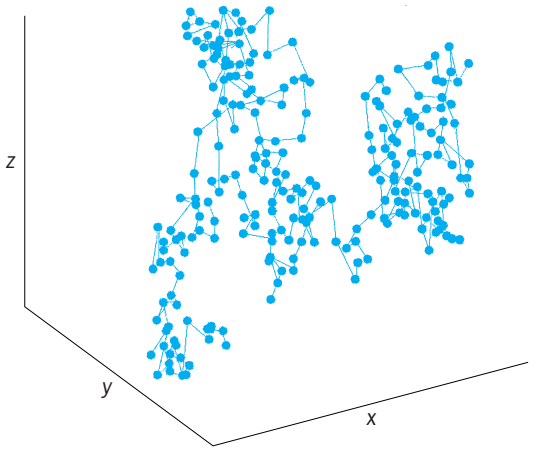

Fig. 3. Brownian motion of a particle in three dimensions.

square root of the molecular weight of the gas. Gases such as $^{235}UF_6$ and $^{238}UF_6$ can therefore be separated by effusion through porous membranes. *See* ISOTOPE SEPARATION.

Electrolyte solutions. Fick's equations also apply to diffusion in electrolyte solutions. The simplest case is a single salt dissolved in a solvent, such as aqueous sodium chloride. The negative and positive ions must diffuse at the same speed to keep all parts of the solution electrically neutral. This means that only one flux is independent and a single diffusion coefficient is sufficient to describe interdiffusion of dissolved salt and solvent.

Interdiffusion coefficients of electrolytes are sensitive to changes in concentration. For a dilute solution of a 1:1 salt, such as NaCl, the interdiffusion is the harmonic average of the diffusion coefficients of the ions: $D = 2D_+ D_-/(D_+ + D_-)$. *See* ELECTROLYTE; ELECTROLYTIC CONDUCTANCE.

Multicomponent interdiffusion. Mixtures of three or more components are of considerable practical importance. Diffusion in these systems differs in important ways from diffusion in simpler two-component mixtures. Adding a new component to an existing mixture changes the rates of diffusion of all the components in the mixture. Also, a new feature appears: coupled diffusion, where the gradient in concentration of each substance causes diffusion of the other substances in the mixture. Coupled diffusion can drive substances from regions of lower to higher concentration, opposite to the expected direction. Diffusion in multicomponent mixtures is described by adding coupled-diffusion terms to Fick's equations.

Coupled diffusion is important in concentrated solutions and in mixed electrolytes because the forces between the diffusing materials are significant. For example, in aqueous solutions of potassium chloride (KCl) and hydrogen chloride (HCl), owing to electrostatic forces, the flux of potassium chloride caused by the gradient in hydrogen chloride can be larger than the flux of potassium chloride caused by its own gradient. Hydrogen chloride diffuses much more rapidly in water-KCl-HCl mixtures than in binary water-HCl mixtures.

Diffusion with convection. In moving fluids, substances are transported by diffusion as well as convection. If the concentration of a substance is C and the solution moves with velocity U_z in the z direction, the flux of substance due to the bulk movement of the solution is CU_z. The net flux is the sum of the contributions from diffusion and convection, as in Eq. (6).

$$J = -D\frac{\partial C}{\partial z} + CU_z \tag{6}$$

See CONVECTION (HEAT).

Dispersion in pipelines. The dispersion of a tracer in a liquid carrier stream is an interesting example of combined diffusive and convective transport. The liquid near the center of the tube moves more quickly than the liquid near the tube wall. This causes an injected tracer to spread out along the axis of the tube. If the flow rate is low enough to prevent turbulence, the combined actions of axial convection and radial diffusion shape the dispersed tracer into a bell-shaped plume. Dispersion is minimized by using short, narrow tubes. Diffusion coefficients of liquids and gases are determined by measuring dispersion profiles. *See* FLUID FLOW.

Ionic diffusion in electric fields. When a salt solution is placed between charged electrodes, electrical forces draw the positive ions toward the negative electrode and the negative ions toward the positive electrode. The flux of an ion caused by an electric field is proportional to the field strength and to the ion's charge, diffusion coefficient, and concentration.

Dividing the drift velocity of a migrating ion by the electric field gives the ion's electric mobility. The electric mobilities of aqueous ions are usually in the range 0.0003–0.001 $cm^2\ s^{-1}\ V^{-1}$. Thus, a 10 V cm^{-1} electric field generates ionic drift velocities of about 0.003–0.01 cm s^{-1}.

Thermal diffusion. If one end of a convection-free column of solution is heated and the other end cooled, solute spontaneously migrates to the cooler regions and solvent migrates to the warmer regions, or vice versa. Diffusion caused by temperature gradients is called thermal diffusion.

Diffusion in nonisothermal solutions is described by adding the term $-CD_T\partial T/\partial z$ to the right-hand side of Eq. (1) to allow for the flux caused by the temperature gradient, giving Eq. (7). The quantity D_T is

$$J = -D\frac{\partial C}{\partial z} - CD_T\frac{\partial T}{\partial z} \tag{7}$$

the thermal diffusion coefficient. Positive values of D_T indicate migration to the cooler parts of a mixture. Substances with negative D_T values migrate to warmer regions.

The accumulation of material at the ends of the solution column produces concentration gradients. Eventually, thermal migration is balanced by ordinary diffusion down the concentration gradient, and a steady state is reached in which there are no further changes in concentration. The ratio of the thermal diffusion coefficient to the interdiffusion coefficient gives the fractional change in solute concentration per degree at steady state, as in Eq. (8). Typical D_T/D

$$\frac{D_T}{D} = -\frac{1}{C}\left(\frac{dC}{dT}\right)_{\text{steady state}} \tag{8}$$

values are about 0.001 K^{-1}, which corresponds to a 1% concentration difference for a 10 K (18°F) temperature difference.

Thermal diffusion is used to purify isotopes and to fractionate polymers, but for most solutions the separations are too small to compete with other techniques. A few solutions have D_T/D values as large as 0.03 K^{-1}. Thus, a 10 K (18°F) temperature difference

can produce a 30% change in concentration, which is not negligible. Thermal diffusion effects can be significant if large temperature gradients are imposed on a solution.

Derek G. Leaist

Solids

Diffusion in solids is an important topic of physical metallurgy and materials science since diffusion processes are ubiquitous in solid matter at elevated temperatures. They play a key role in the kinetics of many microstructural changes that occur during the processing of metals, alloys, ceramics, semiconductors, glasses, and polymers. Typical examples of such changes include nucleation of new phases, diffusive phase transformations, precipitation and dissolution of a second phase, recrystallization, high-temperature creep, and thermal oxidation. Direct technological applications concern diffusion doping during the fabrication of microelectronic devices, solid electrolytes for battery and fuel cells, surface hardening of steels through carburization or nitridation, diffusion bonding, and sintering. *See* CREEP (MATERIALS); FUEL CELL; HEAT TREATMENT (METALLURGY); METAL, MECHANICAL PROPERTIES OF; PHASE TRANSITIONS; PLASTIC DEFORMATION OF METAL; SINTERING; SOLID-STATE BATTERY; SURFACE HARDENING OF STEEL.

The atomic mechanisms of diffusion are closely connected with defects in solids. Point defects such as vacancies and interstitials are the simplest defects and often mediate diffusion in an otherwise perfect crystal. Dislocations, grain boundaries, phase boundaries, and free surfaces are other types of defects in a crystalline solid. They can act as diffusion short circuits because the mobility of atoms along such defects is usually much higher than in the lattice. *See* CRYSTAL DEFECTS.

Diffusion coefficients. The diffusion of atoms through a solid can be described by Fick's equations (1) and (3). For a constant (concentration-independent) diffusivity, solutions of these equations exist for a wide variety of initial and boundary conditions and permit a determination of the diffusion coefficient (D) either from measurements of the concentration distribution as function of the position and time or from measurements of diffusion fluxes and concentration gradients. For example, if a diffusing species is initially concentrated in a thin layer, then after a time t it will have a bell-shaped concentration profile whose width depends on D and t in the same manner as in tracer diffusion.

For diffusion in a material with a chemical composition gradient, solutions of Eq. (3) in closed form are usually not possible. Then numerical methods are often used to deduce so-called interdiffusion coefficients.

Any diffusion coefficient is a material parameter, which depends on the composition, slightly on pressure, and strongly on temperature T. Very often the temperature dependence for a given material is described by an Arrhenius equation (9), where Q is the

$$D = D_0 \exp\left(-\frac{Q}{RT}\right) \tag{9}$$

activation enthalpy, D_0 is a preexponential factor, and R is the gas constant. *See* ENTHALPY.

The most fundamental diffusion process in a solid is self-diffusion and refers to diffusion of A atoms in a solid element A or in an A-B alloy or compound. Following Eq. (9), the self-diffusion coefficients of many metals increase by 7 to 10 orders of magnitude as the absolute temperature is increased from half the melting temperature to the melting temperature. Diffusion of a solute B in a very dilute A-B alloy with the solvent A is denoted as impurity diffusion. Self-diffusion and impurity diffusion in a homogeneous solid can be studied, for example, by utilizing small amounts of radioactive trace atoms. Diffusion in an inhomogeneous solid with a composition gradient is denoted as interdiffusion or sometimes as chemical diffusion.

Crystalline solids differ from gases and liquids in that the symmetry of the atomic arrangement can give rise to different values of D in different directions of the crystal lattice. In isotropic materials such as glasses and polycrystals with random orientation of the individual grains and in cubic single crystals, only one D value characterizes the material. Two diffusion coefficients are necessary for uniaxial (hexagonal, tetragonal, and trigonal) crystals, and three D values characterize diffusion in single crystals of lower symmetry. For example, the D value for diffusion in the basal plane of a hexagonal single crystal is different from the D value perpendicular to this plane. *See* CRYSTAL STRUCTURE.

Mechanisms. Diffusion in crystalline matter takes place by a series of jumps of individual atoms from site to site throughout the crystal. Because of the lattice vibrations, atoms in a crystal oscillate around their equilibrium (substitutional or interstitial) positions with frequencies, ν_0, of the order of the Debye frequency (typically 10^{12}–10^{13} Hz). Usually an atom is confined to a certain site by a potential barrier with barrier height G^M, which corresponds to the Gibbs free-energy difference between the configuration with the jumping atom at the saddlepoint and with that at its equilibrium position. Diffusion is thermally activated, meaning that a fluctuation of thermal energy pushes the atom over the energy barrier which exists between two neighboring sites. From reaction rate theory, the jump frequency ω at which an atomic jump will occur into an empty neighboring site may be written in the form of Eq. (10).

$$\omega = \nu_0 \exp\left(-\frac{G^M}{RT}\right) \tag{10}$$

See BOLTZMANN STATISTICS; FREE ENERGY; LATTICE VIBRATIONS; SPECIFIC HEAT OF SOLIDS; STATISTICAL MECHANICS.

At an atomic level, the diffusivity of (for example, radioactive) tagged atoms in a cubic crystal can be described by Eq. (11), where d is the jump distance

$$D = \frac{1}{6} f d^2 \omega p \tag{11}$$

(some fraction of the lattice parameter), p is the probability that the neighboring site of the jumping atom is empty, and f is the correlation factor, a term which expresses the degree of randomness of successive jumps. Various atomic mechanisms have been identified for bulk diffusion in crystals.

Interstitial mechanism. Solute atoms which are much smaller than the lattice atoms, as for example, hydrogen, carbon, nitrogen, and oxygen in metals, are incorporated in interstitial sites. They diffuse by jumping from interstitial site to interstitial site (**Fig. 4***a*). For a dilute interstitial alloy, the probability p in Eq. (10) is unity. The magnitude of the barrier height G^M is also relatively small, with the result that the diffusion coefficients for interstitial diffusion can be many orders of magnitude larger than those for the self-diffusion of lattice atoms. Interstitial diffusivities near the melting temperature of the solvent can be as high as diffusivities in liquids.

Vacancy mechanism. Self-atoms or substitutional solute atoms migrate by jumping into a neighboring vacant site. Self-diffusion in metals and alloys, and in many ionic crystals and ceramic materials, occurs via the vacancy mechanism (Fig. 4*b*). Attractive or repulsive interactions between the vacancy and substitutional solute atoms may lead to higher or lower diffusivities compared with self-diffusion of the pure solvent. *See* IONIC CRYSTALS.

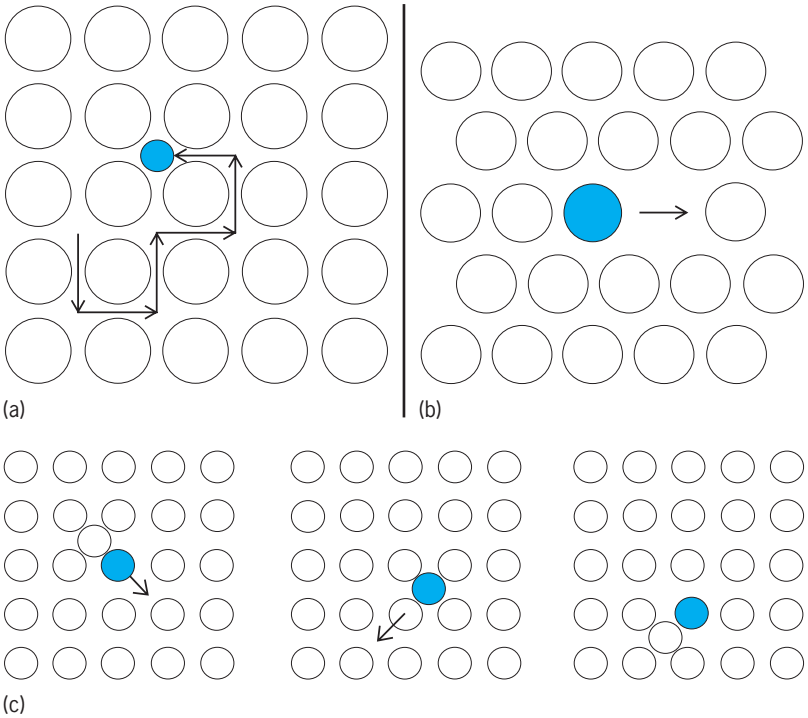

Fig. 4. Diffusion mechanisms. (*a*) Direct interstitial mechanism. (*b*) Vacancy mechanism. (*c*) Self-interstitial mechanism.

In crystals with ionic bonding characteristics such as alkali halides, in ceramics, and in some intermetallic compounds, atoms do not jump between the sublattices of the constituents. This implies that the diffusion coefficients of the various atomic species can differ significantly.

Self-interstitial mechanism. In this case, self-interstitials—extra atoms located between lattice sites—act as diffusion vehicles. A self-interstitial replaces an atom on a substitutional site which then replaces again a neighboring lattice atom (Fig. 4*c*). Self-interstitials are responsible for diffusion in the silver sublattice of silver halides. In silicon, the base material of microelectronic devices, the interstitialcy mechanism dominates self-diffusion and plays a prominent role in the diffusion of many solute atoms, including the most important doping elements. In the close-packed metals, this mechanism is unimportant under thermal equilibrium conditions because of the fairly high formation enthalpy of self-interstitials. It is, however, important for radiation-damage-induced diffusion, where self-interstitials and vacancies are created athermally by irradiation of a crystal with energetic particles. *See* RADIATION DAMAGE TO MATERIALS.

Interstitial-substitutional exchange mechanisms. Some solute atoms (B) may be dissolved on interstitial (B_i) and substitutional sites (B_s) of a solvent crystal (A). For so-called hybrid solutes the diffusivity of B_i is much higher than the diffusivity of B_s, whereas the opposite is true for the solubilities. Under such conditions, the incorporation of B atoms can occur by the fast diffusion of B_i and the subsequent change-over to B_s.

If the change-over involves vacancies (V) according to reaction (12), the mechanism is denoted as the

$$B_i + V \rightleftharpoons B_s \tag{12}$$

dissociative mechanism (sometimes also known as the Frank-Turnbull mechanism or the Longini mechanism). The rapid diffusion of copper in germanium and of some foreign metallic elements in polyvalent metals like lead, titanium, and zirconium has been attributed to this mechanism.

If the change-over involves a self-interstitial of the solvent (A_i) according to reaction (13), the mechanism is denoted as the kick-out mechanism. Examples of this mechanism have been established for some rapidly diffusing foreign elements (for example, gold, platinum, and zinc) in silicon.

$$B_i \rightleftharpoons B_s + A_i \tag{13}$$

Short-circuit diffusion. Any real crystalline material usually contains extended defects such as dislocations, grain boundaries in polycrystals, interphase boundaries in polyphase materials, and free surfaces. These regions are more open structures than the adjoining crystal lattice. The jump frequency of atoms in these regions is much higher and the activation

enthalpy of diffusion lower than in the perfect crystal. Because of the lower activation enthalpy, the mobility of atoms along extended defects decreases less rapidly with decreasing temperature than in the bulk. The result is that at low temperatures (below about half of the melting temperature) diffusion through the lattice becomes slower than diffusion along the extended defects. Then these defects constitute paths of high diffusivity which at low enough temperatures are said to short circuit lattice diffusion.

The rapid diffusion along high-diffusivity paths is often important in surface- and thin-film technologies. In thin-film microelectronic devices, shortcircuit diffusion among various components of the multilayered structures is one of the principal modes of device failure. The fast diffusion character of grain-phase and interphase boundaries also results in the fact that many solid-state reactions at lower temperatures, such as discontinuous precipitation and eutectoid transformations, are controlled by diffusion along boundaries.

Superionic conductors. Superionic conductors, sometimes referred to as fast-ion conductors or solid electrolytes, are solids which have an exceptionally high ionic conductivity over a reasonable temperature range. In some cases the conductivity approaches the magnitude found in molten salts and aqueous solutions of strong electrolytes. Typical values of the conductivity of a superionic conductor are in the range of 0.001 to 10 $(\Omega\,\mathrm{cm})^{-1}$. The conductivity of copper near room temperature is $6 \times 10^5\ (\Omega\,\mathrm{cm})^{-1}$; a typical ionic crystal has a conductivity of $10^{-4}\,(\Omega\mathrm{cm})^{-1}$ close to the melting temperature; an aqueous solution of 0.1 N NaCl has a conductivity of $10^{-2}\ (\Omega\ \mathrm{cm})^{-1}$. Diffusivity D and ionic conductivity σ are related via the Nernst-Einstein equation (14), where k_B denotes Boltzmann's constant,

$$D = \frac{\sigma k_B T}{q^2 N} \tag{14}$$

T absolute temperature, q the ionic charge, and N the volume density of mobile ions. Superionic conduction implies fast diffusion of the ions. Typically a conductivity of 1 $(\Omega\ \mathrm{cm})^{-1}$ corresponds to a diffusivity of about $10^{-5}\ \mathrm{cm}^2\ \mathrm{s}^{-1}$, which is in the same order of magnitude as diffusion in liquids. Because of their high conductivity, superionic conductors are thought of as solid electrolytes with many applications, both realized and potential, for example, in batteries, fuel cells, and sensors. *See* BATTERY; FUEL CELL.

The high ionic conductivity, depending on the material, can be of either cationic or anionic character. As a general rule, superionic conductors are materials with an open structure, which allows the rapid motion of relatively small ions. The most important classes of superionic conductors are the following:

1. Many superionic conductors have a fixed anion lattice and a three-dimensionally disordered cation lattice. Example are those of the silver iodide type (reported as early as 1913), which display a first-order phase transition between a superionic phase at high temperatures and a normal conducting phase at low temperatures. In the case of pure silver iodide (AgI), the high-temperature phase α-AgI exists above 146°C (295°F) with a body-centered cubic sublattice of fixed I^- ions. Each unit cell presents 42 interstitial sites over which the two Ag^+ ions can distribute themselves (**Fig. 5**). Since there are many more sites than Ag^+ ions, the latter can migrate easily. Materials derived from α-AgI with the formula MAg_4I_5, where M = K, Rb, Cs, and NH_4, are superionic at ambient temperature.

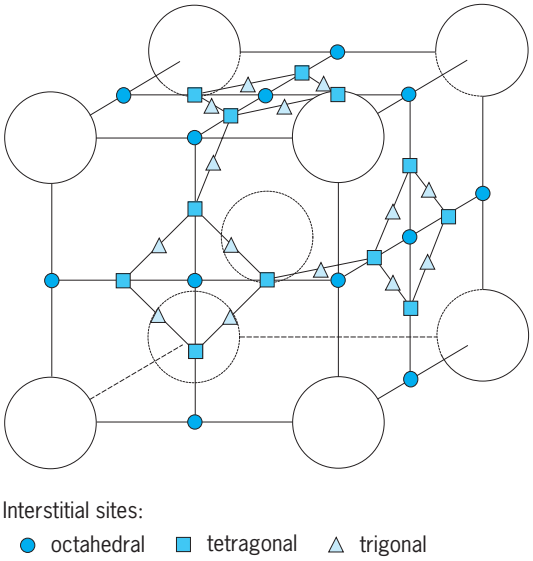

Fig. 5. Crystal structure of α-silver iodide.

2. Some fast-ion conductors have a hexagonal structure in which small cations can move either two-dimensionally in layers or one-dimensionally in channels. Materials in which two-dimensional diffusion occurs are the β-aluminas, which are nonstoichiometric aluminates with the formula $M_2O_nAl_2O_3$, where M = Na, Li, Ag, ... and n ranges from 5 to 11. A material with one-dimensional diffusion is β-eucryptite, an aluminosilicate ($LiAlSiO_4$), in which the lithium ions are located in channels parallel to the hexagonal axis.

3. Zeolites are natural or synthetic hydrated aluminosilicates with an open three-dimensional structure, in which ions are incorporated in cavities of the lattice. *See* ZEOLITE.

4. Superionic conductors are also found among compounds with the fluorite structure, such as some halides ($SrCl_2$, CaF_2, PbF_2) and oxides like ZrO_2 (zirconia), and exhibit fast-ion conductivity. Zirconia is stabilized in the cubic structure by additions of CaO and Y_2O_3.

5. A number of glasses, including lithium borate glass, have been examined as fast-ion conductors.

6. Solutions of ionic salts in certain polymers (polymer electrolytes) can exhibit conductivities in the

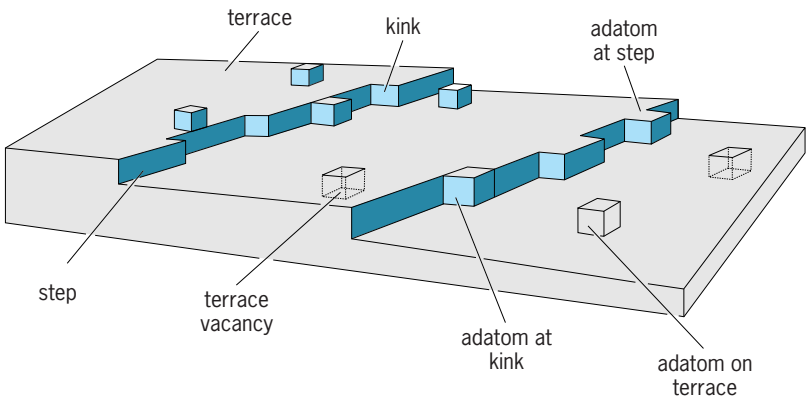

Fig. 6. Model of the surface of a single crystal.

intermediate range. The fact that these materials can be cast as thin, flexible foils led to applications as battery electrolytes.

Surface diffusion. Surface diffusion is the motion of atoms or molecules on top of the surface of a solid material. The diffusivity of an atom on the surface is usually much higher than the diffusivity of the same atom in the interior of the solid, and its activation enthalpy is lower than in the bulk material.

The surface of a crystalline solid has a complicated but well-defined crystallographic structure, which can be described in terms of terraces (low-indexed, vicinal planes), steps, and kinks (**Fig. 6**). Adatoms on the terraces and at steps and surface vacancies occur in thermal equilibrium. More complex defects, such as dislocation emerging points, impurity clusters, high-index facets, and hillocks, may be present. The structure of a real surface will give rise to a mobility of atoms that depends on their position. It will be different, for example, on terraces and on steps. Often an anisotropy of diffusion is also observed, depending on the orientation of vicinal surfaces.

Surface diffusion at temperatures that are not too high occurs via elementary jumps of surface vacancies and adatoms over one atomic distance. When the temperature rises, multiple jumps of adatoms over several atomic distances are possible. Near the melting temperature, even pseudo-liquid behavior has been observed in molecular-dynamic simulations.

Various methods can be used to study surface diffusion. Among them are observations of the migration of single adatoms by means of field-ion microscopy and scanning tunneling microscopy; observations of the redistribution of radiotracer atoms initially deposited on one part of a surface; and topographic methods that monitor the diffusion-induced change in the shape of a solid, which is driven by surface tension.

Radiation-enhanced diffusion. The diffusion coefficient of atoms in solids is dominated by the product of the equilibrium concentration of the diffusion-relevant defects times their mobility. Equilibrium concentration and defect mobility drop steeply with decreasing temperature. As a consequence of this strong temperature dependence, hardly any atomic diffusivity will remain below about 40% of the melting temperature.

In radiation-enhanced diffusion, additional point defects are produced by the incident radiation. High-energy electrons (available, for example, in a high-voltage electron microscope) colliding with the lattice atoms can produce pairs of vacancies and self-interstitials (Frenkel pairs). Irradiation with heavier particles, such as protons, fast neutrons, alpha particles, and heavy ions, causes radiation damage to the crystal lattice which introduces many Frenkel pairs per incident particle. Therefore, not only will the usual diffusion path be enhanced, but also new channels can be opened by way of defects (such as self-interstitials in the case of metals) which are not available in normal thermally activated diffusion. Radiation-enhanced diffusion can occur at relatively low temperatures provided that thermal activation permits sufficient mobility of at least one of the defect species produced during irradiation. Radiation-enhanced diffusion is of importance in materials science since it affects the properties of materials exposed to particle irradiation. Helmut Mehrer

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Digenea

A group of parasitic flatworms or flukes constituting a subclass or order of the class Trematoda, in the phylum Platyhelminthes. The name Digenea refers to the two types of generations in the life cycle: (1) the germinal sacs which parasitize the intermediate host (a mollusk or rarely an annelid) and reproduce asexually; and (2) the adult which is primarily endoparasitic in vertebrates and reproduces sexually. The adult usually is hermaphroditic, but many of the blood flukes and a few others are dioecious. Vertebrates of all classes, except the Cyclostomata, serve as definitive hosts. Those feeding on aquatic plants and animals harbor the greatest variety of digenetic trematodes, but several species occur in strictly terrestrial hosts. Effects on the vertebrate vary from no apparent harm to severe and even fatal injury. Control of flukes is by preventive measures, including interruption of their life cycles and treatment with drugs. *See* TREMATODA.

Life histories of digenetic trematodes. The life histories are complex and unlike those of other trematodes in which the egg gives rise to a single adult (**Fig. 1**). Instead, the miracidium, which develops from the egg, enters the intermediate host and becomes a germinal sac, the miracidial or mother sporocyst. Its germinal cells give rise to a second, or daughter, sporocyst generation or to one of rediae, which differ from sporocysts only in having a pharynx and saclike gut. The redia, which may or may not produce a second generation of rediae, may be present or absent in the life history. Germinal cells of the definitive sporocyst or redial generation develop into cercariae. The miracidial sporocyst rarely produces cercariae directly, and instances are known in which a redia or sporocyst may give rise to both its own kind and also to cercariae, thus maintaining the infection indefinitely in the intermediate host after the production of cercariae begins. The mode of reproduction in sporocysts and rediae has been disputed. It has been interpreted as diploid parthenogenesis in recent studies. The number of cercariae produced by the germinal sacs resulting from a single miracidium varies from a few dozen to thousands or even millions.

The cercaria usually leaves the intermediate host and then encysts, either on vegetation or after penetrating a more or less specific second intermediate, or vector, host. The encysted larva, called the metacercaria, is eaten by the vertebrate and becomes the adult fluke in that host. Cercariae of blood flukes enter the vertebrate directly by penetrating the skin. Swimmer's itch in the United States results when such larvae of avian blood flukes penetrate the human skin and die because humans are unsuitable hosts. Fish acquire certain trematodes by ingesting cercariae, which are large and attract attention. Cercariae rarely remain in the molluscan host, with or without encysting, and are eaten with the host by the vertebrate. *See* CERCARIA.

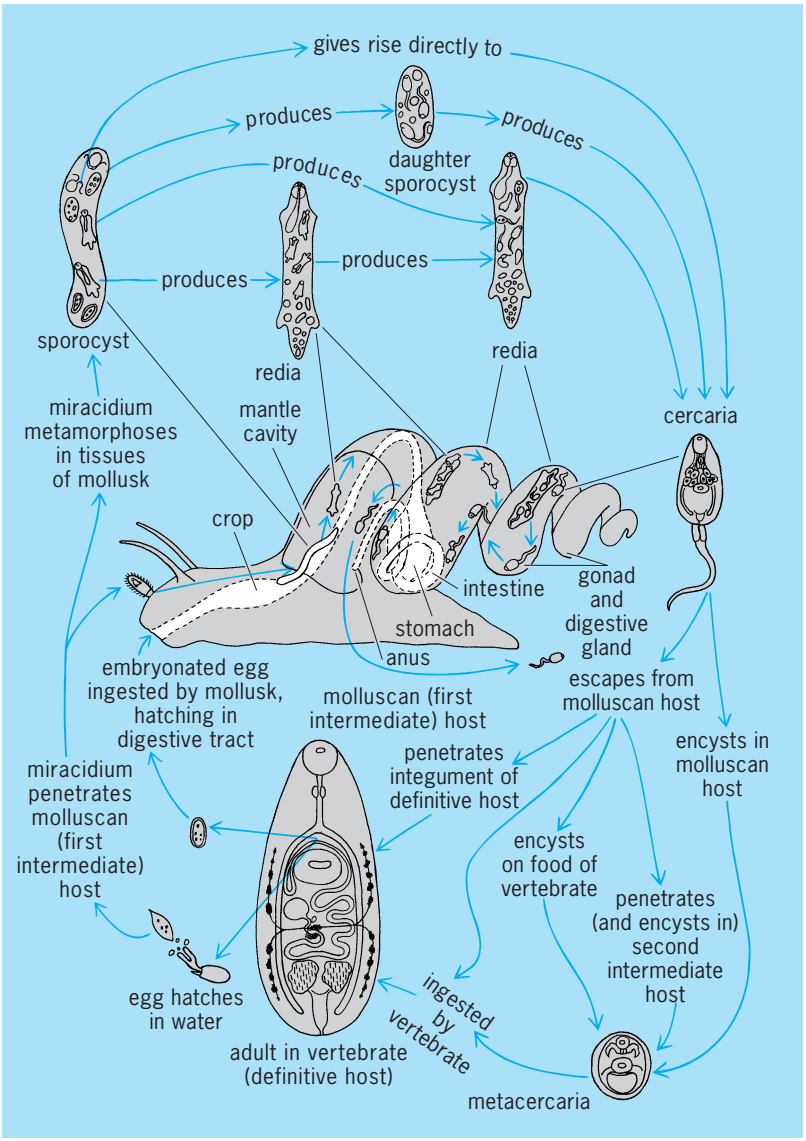

Fig. 1. Digenetic trematode cycle. (*After R. M. Cable, An Illustrated Laboratory Manual of Parasitology, Burgess, 1940*)

Morphology. The adult differs from other trematodes in several respects, the most obvious being the absence of a posterior adhesive organ (opisthaptor) with sclerotized hooks, plates, or multiple suctorial structures. Usually the mouth is anterior and opens into an oral sucker, but in the gasterostomes it is ventral and without a sucker (**Fig. 2**). A second sucker, the ventral one, or acetabulum, is at the posterior end of amphistomes, on the ventral surface of distomes, or absent in monostomes. The entire ventral surface may be concave and serve for adhesion, with or without the aid of a ventral sucker and glandular structures, or a portion of that surface may form a complex adhesive organ, as in the holostomes. Adult digenetic trematodes vary from about 0.2 mm (0.008 in.) to well over 1 m (3.28 ft) in length, but usually are less than 2.5 cm (1 in.) long. They occur in any part of the vertebrate that provides egress for their eggs and also in the circulatory system, from which eggs work through the tissues and escape in the feces or urine.

Classification. There is no general agreement concerning the classification of the Digenea. When few life histories were known, the group was divided into orders and families on the basis of such adult features as the number and position of suckers and components and arrangement of the reproductive system. Instances of divergent and convergent evolution with respect to those features have been revealed by life history studies which accordingly have prompted drastic revision of taxonomy of the Digenea, especially at the superfamily and order levels. The scheme of G. La Rue proposed two superorders, the Anepitheliocystidia and Epitheliocystidia, based on the presence or absence of an epithelial wall of the excretory bladder as revealed by studies on cercariae. Each superorder is divided into orders on the basis of the cercarial tail, the relationship of the excretory system to it, and the structure of the miracidium. The cercarial type is the basis of the superfamily, with that category being divided into families largely according to adult structure. In that respect, trematodes of the subclass or order Aspidobothrea

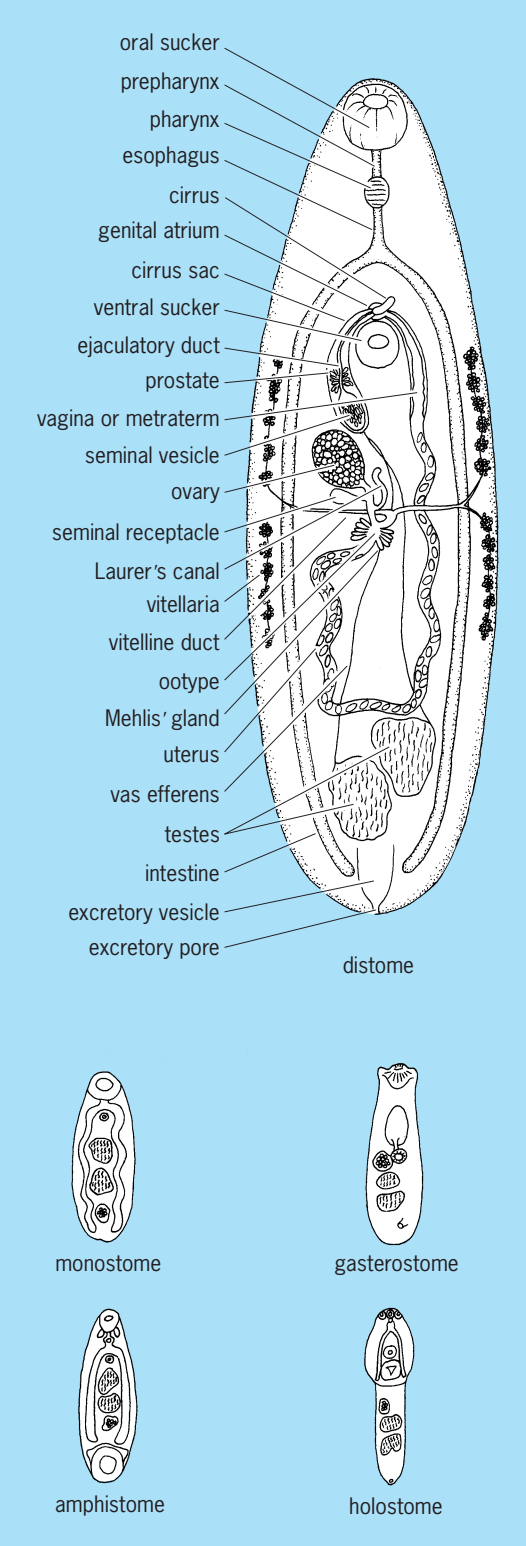

Fig. 2. Digenetic trematodes. Diagrams of various adult types. (*After R. M. Cable, An Illustrated Laboratory Manual of Parasitology, Burgess, 1940*)

(Aspidogastrea) resemble the Digenea more than the Monogenea, but so far as is known do not have germinal sacs and larval reproduction in an intermediate host. *See* ASPIDOGASTREA; PLATYHELMINTHES.

Raymond M. Cable

Digestion (invertebrate)

The process by which the complex organic molecules in foods are broken down into simpler units which can be utilized by cells for energy or incorporated into living protoplasm. Frequently, there is an essential mechanical pretreatment of the food, either before or after it is taken into the body. Mechanical treatment is especially needed in herbivorous animals that depend on the utilization of cellulose and other plant materials, or where the prey has a hard exoskeleton. Many animals have devices for crushing, grinding, or mixing.

The final processes of digestion are always chemical. Reactions in chemical digestion commonly involve hydrolysis, the accelerated addition of water to the food molecules resulting in their cleavage to simpler units. The water brings about the decomposition, and the hydrolytic digestive enzymes accelerate the process.

Enzymes. Digestive enzymes are highly specific in their action, each acting upon a particular type of molecule or even upon a particular type of bond within a molecule. Enzymes have a hydrogen ion concentration at which activity is maximal (pH optimum). The pH values in invertebrate digestive chambers generally fall in the range of pH 5 to 8, where invertebrate digestive enzymes are active. A more acid or alkaline environment is produced in special cases.

Few invertebrates have been examined thoroughly as to the number and chemical nature of enzymes involved in digestion. A complete digestive enzyme complement includes proteinases, carbohydrases, esterases including lipase, and nucleases. Proteinases hydrolyze proteins by breaking peptide bonds between amino acids. Some proteinases (enopeptidases) hydrolyze bonds within the chain of amino acids, while others (exopeptidases) act only on the bond joining the terminal amino acid to the main chain. Both categories include several enzymes specific for particular amino acids. By their combined actions, the protein is digested to individual amino acids. Similarly, a carbohydrase such as amylase hydrolyzes polysaccarides such as starch or glycogen to maltose (a disaccharide) and glucose (a monosaccharide). Various disaccharidases have been shown to be widely distributed in vertebrates. Some esterases are nonspecific, whereas lipase hydrolyzes neutral fats and oils with long-chain fatty acids. Nucleases hydrolyze DNA and RNA to nucleotides. The complement of digestive hydrolases may also include phosphatases and sulfatases.

The presence of some enzymes is related to specific diets. Phytophagous animals may have cellulase or pectinase or both for the digestion of components in the plant wall. Chitinase activity is often reported in invertebrates that ingest fungi or animals with chitinous integuments. Activities of these enzymes may often be due to enteric bacteria or protozoan symbionts. However, the production of collagenase for the digestion of vertebrate connective tissue by some maggots seems well established, and the larvae

of some intestinal parasitic nematodes may also produce it. *See* ENZYME.

Intracellular and extracellular. Digestion takes place within a chamber which is sometimes a cytoplasmic vacuole formed by endocytosis at the cell surface. This is intracellular digestion (**Fig. 1**). Single-unit-membrane bodies called lysosomes are formed in the Golgi region of the cell. They contain a variety of hydrolases operating most effectively in an acid medium. The lysosomes fuse with the food vacuoles. This process is considered to be a primitive method of digestion and is commonly found in protozoa and sponges. But it is also prevalent in the coelenterates, platyhelminthes, annelids, mollusks, and minor phyla in which immobile or mobile phagocytic cells may be involved. *See* ENDOCYTOSIS; LYSOSOME.

In most metazoans it is also possible for digestion to take place outside cells in a chamber such as a gastrovascular cavity or an alimentary canal (digestive tract). This is called luminal extracellular digestion. Species are found in most phyla that demonstrate either partially or wholly luminal extracellular digestion. Some animals demonstrate extracorporeal extracellular digestion, in which digestive enzymes are released to the outside of the animal. The digested material is taken up orally or through the surface of the organism. Even some protozoa may demonstrate this phenomenon. In metazoans demonstrating luminal extracellular digestion, it is often difficult to determine whether the digestive enzymes are exogenous (food or bacterial) or endogenous (glandular) in origin. Further, the enzymes may originate from intracellular food vacuoles at the time of egestion of indigestible residues.

Some animals have adapted to a highly specialized feeding habit and have abandoned the biosynthesis of some digestive hydrolases. Endoparasites often represent the extreme of such adaptations. Frequently a digestive tract is absent, and no evidence for the use of extracorporeal digestive enzymes can be found.

Digestive tract. The typical invertebrate digestive tract consists of an ectodermal foregut and hindgut and an endodermal midgut which is usually the major site of digestion. Particular regions of the foregut may be recognized and given special names, such as buccal cavity, pharynx, and esophagus; and the esophagus may be further differentiated into a crop or gizzard. The midgut may contain a stomach along with the intestine; and the hindgut consists of a colon or a rectum or both. It should be noted that the above regions, as designated in invertebrate digestive tracts, do not necessarily correspond in structure or function to the regions in vertebrate tracts with the same names.

Extracellular digestive enzymes are secreted by unicellular glands in the epithelium or by glandular diverticula. Mucous cells are widely distributed for lubrication. The midgut epithelium in some animals secretes a membrane between it and the contents of the gut. This peritrophic membrane is protective against abrasion and may act as a filter. The presence of a rotating strand or rod for food movement and sorting is common in microphagous metazoans, especially ciliary feeders.

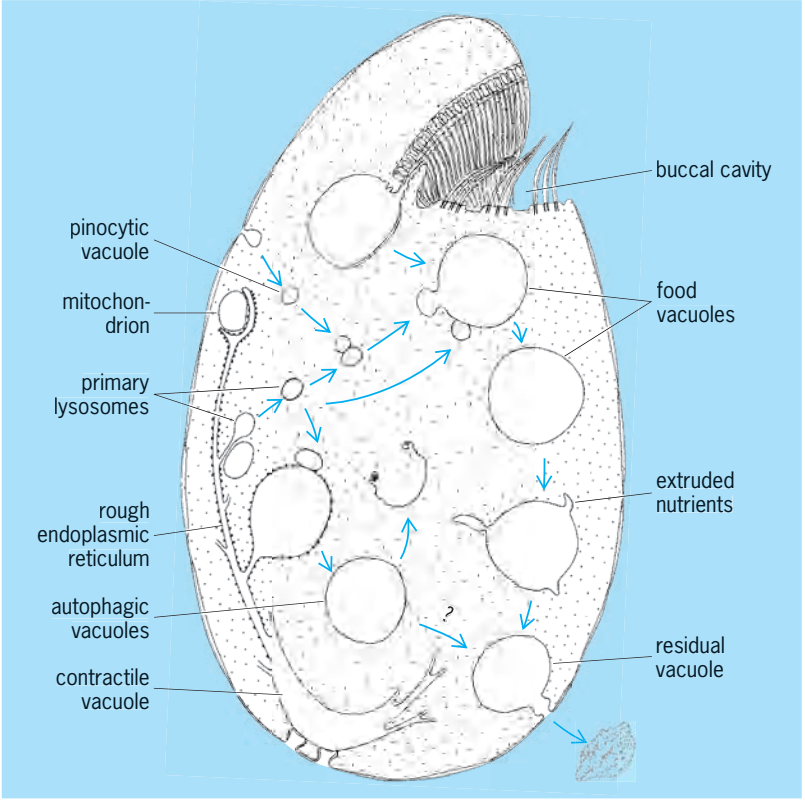

Fig. 1. Possible structural pathways of intracellular digestion in *Tetrahymena pyriformis*. (*After A. M. Elliott and G. L. Clemmons, J. Protozool., 13(2):311–323, 1966*)

Protozoa. In most protozoa, phagocytic food vacuoles and pinocytotic vesicles are involved in food intake (Fig. 1). The organic components of the foods are digested by the activities of the acid hydrolases contributed by the coalescing lysosomes. The vacuole becomes alkaline, and bacterial enzymes as well as those in the engulfed food may also contribute to the digestive process. Many protozoan species, including free-living and parasitic forms, release hydrolases into the medium. The intracellular and extracellular enzymes may exhibit different properties, and the origin of the latter is still an open question. The properties of protozoan extracellular enzymes may differ from luminal enzymes of metazoan animals. The significance of enzyme secretion by free-living forms is unclear. In the parasitic forms, these enzymes may enhance the invasiveness of the organisms. *See* PROTOZOA.

Porifera. Digestion in sponges appears to be entirely intracellular. Most of the cells in contact with the medium are able to engulf food particles, but the flagellated collar cells (choanocytes) are the protagonists. Bacteria and other food particles captured by the choanocytes are transferred to amebocytes, where they are digested. *See* PORIFERA.

Cnidaria. In cnidarians (coelenterates), luminal digestion occurs in a gastrovascular cavity in both polyp and medusa types. Intracellular digestion occurs by phagocytic activity in gastrodermal cells.

Corals remove amino acids directly from the external medium. In sea anemones, enzymes are secreted from the gastric filaments. The saclike mass of mesenteric filaments adheres closely to the surface of the food. One of the proteinases has properties remarkably similar to vertebrate chymotrypsin and may be produced as an inactive zymogen. Tryptic activity is also present. These enzymes are also found in the sea pansy. These findings suggest that divergence of the common ancestors of these enzymes took place before the evolution of cnidarians. In sea walnuts (Ctenophora), luminal digestion begins in the pharynx, and intracellular digestion is completed in the stomach and canal system. Indigestible wastes are passed out through an anal pore. *See* CNIDARIA.

Platyhelminthes. Among turbellarian flatworms, acoels employ a syncytium of cells to ingest food. Vacuoles formed by the syncytium act as temporary gut lumina, and digestion proceeds rapidly within them. In rhabdocoels and probably alloeocoels, club-shaped gland cells secrete endopeptidases into an acid intestine prior to phagocytosis. In triclads, pharyngeal gland cells and intestinal sphere cells secrete endopeptidases prior to intracellular digestion. Digestion may be entirely luminal in polyclads. In flukes (Trematoda), proteolytic secretions produced in the glandular portion of the modified pharynx may partially digest host tissue extracorporeally before luminal digestion proceeds. There is some evidence that intracellular digestion follows. In blood feeders, gland cells in the oral sucker may secrete enzymes which cause histolysis of host tissue and initiate luminal digestion when food finally reaches the gut. Luminal enzymes may be bound to the membranes of absorptive cells. An intracellular proteinase digests globin at pH 3.9. There may also be direct uptake of nutrients through the body surface. Tapeworms (Cestoda) have no alimentary system and must use direct uptake. *See* PLATYHELMINTHES.

Rhynchocoela. The development of an anus in the digestive system of ribbon worms marks a significant departure from the digestive system of flatworms. This system is a much more efficient arrangement and one that is present in almost all higher invertebrates. In some nemerteans there is an initial extracorporeal phase of digestion, and epidermal enzymes may have a role in the uptake of nutrients from the enviroment. In others there is an initial acid-secreting mechanism in the acidophilic gland cells of the foregut which kills the prey. The intestine is richly supplied with gland cells. Digestion is initially luminal but is concluded intracellularly in phagocytic cells. *See* RHYNCHOCOELA.

Pseudocoelomates. Limited information is available concerning digestion among the pseudocoelomate animals (rotifers, gastrotrichs, kinorhynchs, gnathostomulids, entoprocts). Glandular areas of the digestive tract have been identified, and digestion is believed to be primarily if not completely luminal. The free-living adults of the hairworms (Nematomorpha) lack a digestive tract and do not feed. The parasitic juveniles absorb food directly. The spiny-headed worms (Acanthocephala) have no digestive tract and absorb nutrients directly through the body surface. The enzymes from the pharyngeal glands of roundworms (Nemata) may contribute to extracorporeal digestion or combine with additional luminal enzymes secreted from the anterior intestine. *See* ENTOPROCTA; GASTROTRICHA; GNATHOSTOMULIDA; KINORHYNCHA; ROTIFERA.

Mollusca. The molluscan alimentary canal is a highly organized ciliated surface designed for continuous processing of material. The stomach and associated diverticula have complex grooves and ridges that act as ciliary sorting tracts. Primitive gastropods and protobranchiate bivalves have salivary glands which secrete mucus that entangles the ingested food particles (**Fig. 2***a*). The anterior stomach is lined with chitin except for the ciliated sorting region which receives the ducts of the digestive glands. The posterior stomach consists of a ciliated style sac in which ciliary rotation of the ingested mucous mass produces a mucous food string drawn through the esophagus and stomach. Small food particles from the rotating mucous mass (protostyle) are sorted and conveyed to the digestive glands for intracellular digestion. Large particles are carried to the intestine for fecal pellet formation. Luminal digestion occurs in the stomach from enzymes secreted by the digestive glands. A crystalline style in a style sac completely separated from the stomach is characteristic of many bivalves and filter-feeding snails (Fig. 2*b*). The crystalline style is a long compact rod involved in chemical and mechanical digestion. The style sac secretes additions to the style base as abrasion of the protruding anterior style on the chitinous gastric shield releases style enzymes, including amylase and possibly lipase and cellulase. The presence of mucoprotein in the crystalline style precludes the extracellular digestion of proteins. A fold (typhlosole) regulates flow into and out of the main ducts of the digestive glands.

The squid shows the greatest modification of the basic pattern of the molluscan digestive tract (Fig. 2*c*). The midgut gland is divided into a large distal liver and a small proximal pancreas. A large thin-walled cecum, originating in the proximal stomach, receives the main duct of the midgut and opens into the intestine. The adaptation facilitates rapid digestion. The digestive enzymes are produced exclusively in the digestive gland. The pancreatic region secretes enzymes into the cecum, and the contractions of the cecal sac drive them into the stomach. When the contents of the stomach become liquid, the cecal valve opens, and the fluid is passed into the cecum. The hepatic portion of the digestive gland then secretes enzymes into the cecum. Digested material is absorbed in both the cecum and the intestine. *See* MOLLUSCA.

Arthropoda. Extracorporeal secretion of enzymes is a common initial phase of digestion in many arachnids, including scorpions, mites, and spiders. The epithelium of the stomach diverticula may contain zymogen cells for luminal digestion and cells for intracellular digestion in the final stages. Many arachnids survive long periods (1–2 years) without feeding.

The stomach of decapod crustaceans is the site of mechanical trituration and filtration of food particles. Enzymes pass from the digestive gland, and luminal digestion takes place in the stomach. Soluble products flow back into the digestive gland.

The chemical natures of the digestive enzyme secretions of decapod crustaceans and insects have been elucidated more than enzymes from any other group of invertebrates. Most insects have a pair of salivary glands which are involved in luminal carbohydrate digestion, but the midgut (ventriculus or stomach) is the main site of enzyme production and secretion. A proventriculus may act as a macerating gizzard (for solid food) or act as a valve in fluid feeders. Numerous symbiotic microorganisms in the midgut ceca may be major contributors to the digestive process. This is probably the case in the digestion of beeswax by a larval moth. In contrast, a silverfish that feeds on bark has been shown unequivocally to produce its own (endogenous) cellulase. Another special adaptation in insects is the high pH (10) obtained in some moths, beetles, and lice, which alters the keratin of fur, wool, and feathers for subsequent hydrolysis by the normal complement of proteinases. The stable fly and housefly have a midgut proteinase with a very low pH optimum (1.5–2.5) for an invertebrate digestive enzyme. Some predatory flies and true bugs use their salivary secretions for extracorporeal digestion. The walking worms (Onychophora) have similar extracorporeal digestion mechanisms. *See* ARTHROPODA; ONYCHOPHORA.

Annelida. In general, no specialized digestive glands are used by wandering marine polychaete worms (Annelida). The pharynx, esophagus, or anterior intestine may secrete enzymes, and coelomocytes and amebocytes may take part in intracellular digestion. In most sedentary polychaetes, luminal digestion and some absorption occur in the stomach. Polychaetes are able to accumulate amino acids, fatty acids, and glucose through the body wall directly from dilute solutions. Proteinases produced externally by stationary tubiculous polychaetes not only may act in keeping the tube free of attaching organisms but may have a role in extracorporeal breakdown of proteins.

The alimentary canals of terrestrial oligochaetes differ from those of aquatic worms. A broad range of enzymes are secreted for luminal digestion, but the origin of specific enzymes is often unknown. In bloodsucking leeches, unicellular salivary glands secrete a proteinase, hyaluronidase, and hirudin (histamine) extracorporeally. Sucked blood is stored in large crop diverticula for up to 200 days, and the mechanism of luminal digestion is still unclear. In contrast, digestion is rapid in predatory species. *See* ANNELIDA.

Sipuncula, Echiurida, and Priapulida. In peanut worms (Sipuncula) the descending coil of a long intestine is the apparent source of enzymes, but luminal digestion predominates in the loop area. In echiurids the siphonal intestine contains a digestive fluid with a complement of hydrolases. They have also been shown to accumulate amino acids directly from dilute solutions. In priapulids the intestinal fluid is alkaline (pH 8.3) and contains tryptic and chymotryptic activities. *See* ECHIURIDA; PRIAPULIDA; SIPUNCULA.

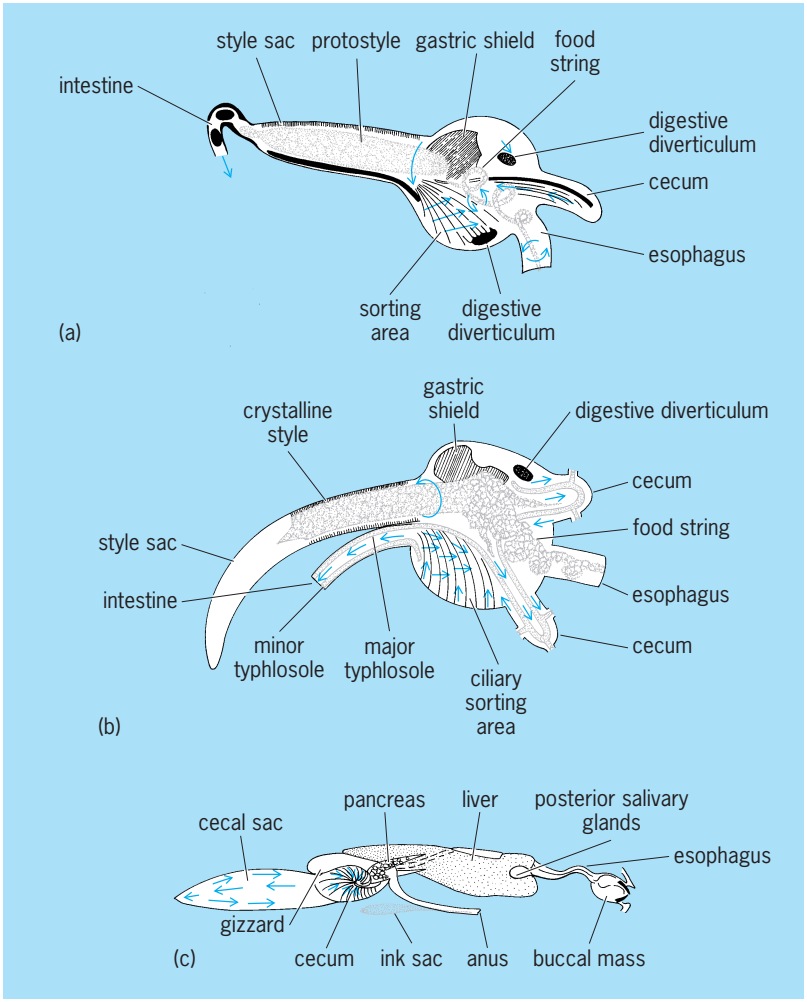

Fig. 2. Molluscan digestive systems. (*a*) Idealized form of the stomach in an early prosobranch. (*b*) A eulamellibranch stomach. (*c*) *Loligo* alimentary canal. (*After J. E. Morton, Molluscs, 4th ed., Hutchinson, 1967*)

Ectoprocta and Phoronida. Limited information indicates most digestion is extracellular in ectoprocts. Fat is digested intracellularly in some marine species. Digestion by brachiopods is chiefly intracellular within the digestive gland, but carbohydrate digestion may be extracellular. Phoronid digestion is probably intracellular within the stomach. *See* PHORONIDA.

Echinodermata. Among the Echinodermata, little is known about digestion in the crinoids and in basket and brittle stars. Coelomocytes appear to play a major role in sea cucumbers; they may invade the gut lumen to release enzymes. The esophagus and intestine are probably the digestive sites in sea urchins, and the cecum may secrete the enzymes. Coelomocytes migrate to the external surface of the urchin to predigest a variety of materials which are then ingested. The pyloric ceca of the starfish secrete enzymes in the cardiac stomach, where luminal digestion occurs. In some, the stomach may be

everted so that digestion begins outside the animal. Endopeptidases from the ceca, such as trypsin and carboxypeptidase, have been comprehensively studied. *See* ECHINODERMATA.

Chordata. Digestion is probably extracellular in sea squirts (Urochordata) and within the stomach, which contains abundant secretory cells. In the lancelet (Cephalochordata), mechanical digestion is similar to that seen in many mollusks. The mixing and sorting take place in the midgut by ciliary action. Small particles from the rotating mass are digested intracellularly in a blind digestive cecum. *See* CHORDATA; FEEDING MECHANISMS (INVERTEBRATE).

Edward J. Devillez

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Digestive system

The vertebrate digestive system consists of the digestive tract and ancillary organs that serve for the acquisition of food and assimilation of nutrients required for energy, growth, maintenance, and reproduction. Food is ingested, reduced to particles, mixed with digestive fluids and enzymes, and propelled through the digestive tract. Enzymes produced by the host animal and microbes indigenous to the digestive tract destroy harmful agents and convert food into a limited number of nutrients, which are selectively absorbed. The digestive systems of vertebrates show numerous structural and functional adaptations to their diet, habitat, and other characteristics. Carnivores, which feed exclusively on other animals, and species that feed on plant concentrates (seeds, fruit, nectar, pollen) tend to have the shortest and simplest digestive tract. The digestive tract tends to be more complex in omnivores, which feed on both plants and animals, and most complex in herbivores, which feed principally on the fibrous portions of plants.

Gut structure and function can also vary with the habitat and other physiological characteristics of a species. The digestive tract of fish has adaptations for a marine or fresh-water environment. The hindgut of terrestrial vertebrates plays an important role in the conservation of the fluids, electrolytes, and nitrogen in the digestive secretions, and in the urinary excretions of reptiles and birds. Fish, amphibians, and reptiles are ectotherms, whose body temperature, metabolic rate, and nutritional requirements vary with their ambient temperature. The endothermic birds and mammals require a higher and more continuous input of nutrients for the maintenance of a constant body temperature. The basal metabolic rate per gram of body weight increases with a decrease in the body mass. Therefore, small animals must process larger amounts of food per gram of body weight, thus limiting their maximum gut capacity and digesta retention time.

Anatomy. Because of wide species variations, the digestive system of vertebrates is best described in terms of the headgut, foregut, midgut, pancreas, biliary system, and hindgut.

Headgut. The headgut consists of the mouthparts and pharynx, which serve for the procurement and the initial preparation and swallowing of food. Movable, articulated jaws are found in all vertebrates except the cyclostomes (lampreys and hagfish), which are primitive, parasitic fish. Teeth, used for grasping, cutting, or tearing of food, are located in the jaws or other mouthparts of most fish. However, in some species, such as the chub, they are located in the pharynx and used as a grinding mill (**Fig. 1**), while the mouthparts of basking sharks and paddlefish are used for the filtering or sorting of food. Complex mechanisms for the microfiltration of ingested food are seen also in some species of larval amphibians, as well as flamingos and baleen whales. Some larval amphibians have horny teeth that are used for procuring prey or grazing on plants, but all adult amphibians are carnivores with a weak dentition that is used only for grasping and positioning prey. Most reptiles have teeth that are used for the procurement, cutting, tearing, or crushing of food, but turtles and birds use a beak for these purposes.

A few mammals have no teeth. The ant and termite eaters in five mammalian orders have simple teeth and weak jaws. However, the teeth of most mammals consist of incisors and canines for grasping, cutting, and tearing, and premolars and molars with uneven occluding surfaces. The articulation and musculature of the mammalian jaws allows both the vertical movement of the lower jaw and either its lateral movement, as seen in most species, or the forward and backward movement, as seen in rodents and elephants. Thus, mammals have the unique ability to use their premolar and molar teeth for the grinding as well as the crushing of food. A movable tongue aids in the procurement of food by frogs, toads, woodpeckers, and mammalian anteaters, and for placement of food between the molars of mammals. Oral glands secrete adhesive materials in frogs, toads, woodpeckers, and mammalian anteaters, toxins in some reptiles, and fluids that aid in the deglutition or swallowing of food in reptiles, birds, and mammals. *See* ORAL GLANDS; TONGUE.

Foregut. The foregut consists of an esophagus for the swallowing of food and, in most species, a stomach that serves for its storage and initial stages of digestion. The esophagus of most vertebrates is lined with a multilayer of cells that are impermeable to absorption. In most birds it contains the crop, an outpocketing of its wall that provides for the temporary storage of food. A stomach is present in all but the cyclostomes and some species of advanced fish (Fig. 1) and in the larval amphibians. In most vertebrates it consists of a dilated segment of the gut that is separated from the esophagus and midgut by muscular sphincters or valves. This is often referred to as a simple stomach. However, in birds these functions are carried out by the crop (storage), proventriculus

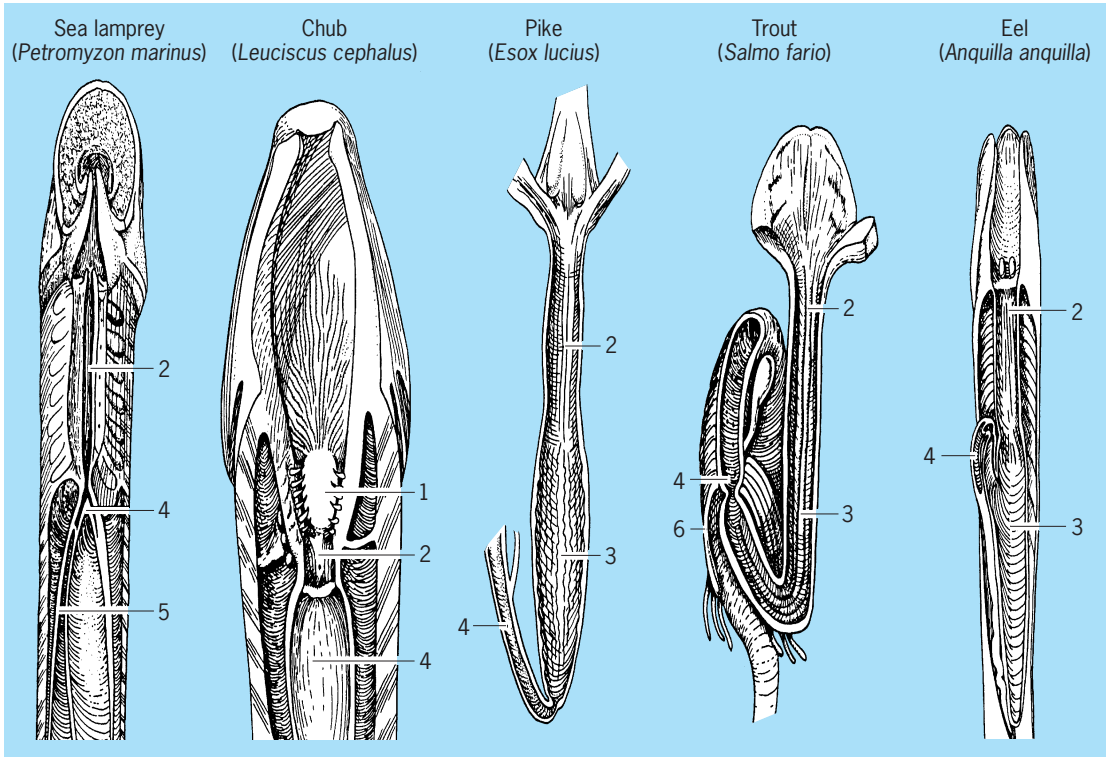

Fig. 1. Major variations in the foregut and midgut of fish. The teeth of most fish are located on the jaws or other mouthparts, but their location in the pharynx (1) of some species such as the chub allows mastication of food without the loss of fine particles through the gills. The esophagus (2) varies in length. The stomach (3) is absent in cyclostomes, such as the lamprey, and in some advanced species, such as the chub. Where present, it may be straight as in the pike, U-shaped like that of the trout, or Y-shaped with a gastric cecum, as seen in the eel. The intestine (4) tends to be relatively short with little distinction between the midgut and hindgut in most fish, but the lumen surface area and digesta retention time of the midgut are increased by a spiral valve (5) or pyloric ceca (6) in some species. (*Modified from W. Harder, Anatomy of Fishes, Part II, Schweizerbart'sche Verlagsbuchhandlung, 1975*)

(secretion), and gizzard (grinding or mastication). *See* ESOPHAGUS; STOMACH.

In most vertebrates, a major portion of the stomach is lined with a proper gastric mucosa (epithelium), which secretes mucus, hydrochloric acid (HCl), and pepsinogen. The distal (pyloric) part of the stomach secretes mucus and bicarbonate ions (HCO^-_3), and its muscular contractions help reduce the size of food particles and transfer partially digested food into the midgut. The stomach of reptiles and most mammals has an additional area of cardiac mucosa near its entrance, which also secretes mucus and bicarbonate ions.

The stomach of some mammals differs from that of other vertebrates with respect to its epithelial lining, structure, or size (**Fig. 2**). In some of the species belonging to 10 of the 20 mammalian orders, the cardiac mucosa is preceded by a region of stratified squamous epithelium, similar to that of the esophagus. This can vary from a narrow region, as seen in the pig, to one-half of the stomach, like that of the horse and rat, to a major fraction of the stomach, as seen in the ox and llama, to the entire stomach of monotremes (spiny anteaters and platypuses). The presence of this region in the stomach of anteaters and many herbivores suggests that it serves as protection against physical damage. Eight mammalian orders include species with an enlarged stomach, which may be divided into permanent compartments, as seen in the rat, ox, and llama.

Midgut. The midgut or small intestine is the principal site for the digestion of food and the absorption of nutrients. It is lined with a single layer of cells that secrete mucus and fluids, contain enzymes that aid in the final stages of carbohydrates and protein digestion, and absorb nutrients from the lumen into the circulatory system. The surface area of the lumen can be increased by a variety of means.

The lumen surface of the midgut of fish is increased by folds that form a spiral valve in some species and by 1–1000 pyloric ceca (blind sacs) in many other species (Fig. 1). In higher vertebrates the lumen surface is increased by the presence of villi, which are macroscopic projections of the epithelial and subepithelial tissue (**Fig. 3***a*). A crypt at the base of each villus contains cells that secrete fluid into the gut and hormones that help regulate the functions of the digestive system. The crypts also contain cells that serve as precursors for the generation of absorptive and mucus-producing cells, which migrate up the villus surface and are eventually sloughed off into the lumen.

The lumen surface is also expanded by a brush border of microvilli on the lumen-facing (apical) surface of the midgut absorptive cells in all vertebrates (Fig. 3*b*). The brush border membranes contain

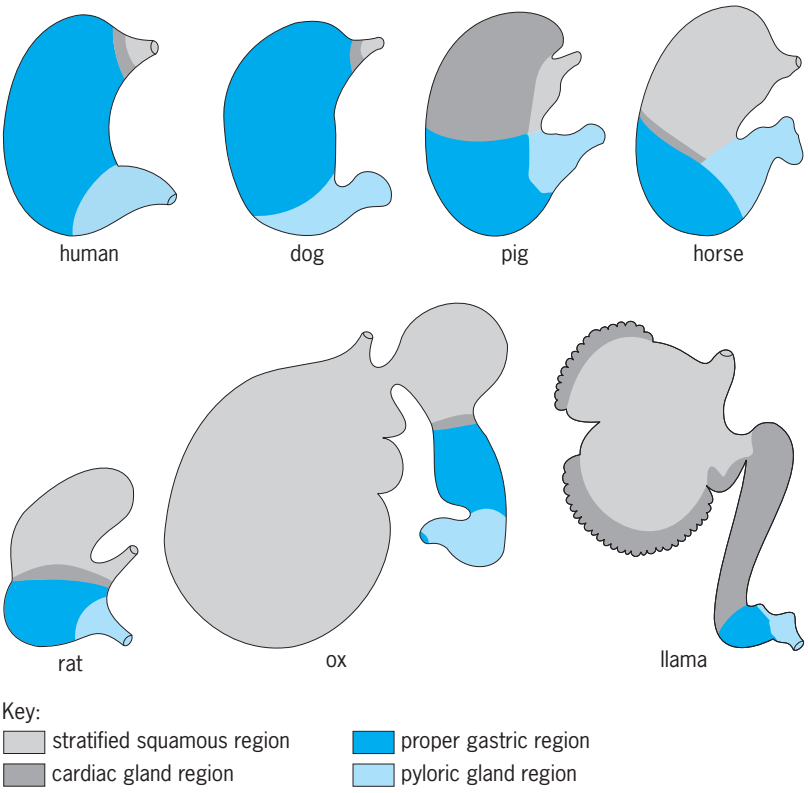

Fig. 2. Variations in the distribution of gastric epithelium and the degree of compartmentalization in the stomach of mammals. Stomachs are not drawn to the same scale; the stomach contents of the ox are approximately 70 times the volume and 14 times the weight of human stomach contents. (*After C. E. Stevens and I. D. Hume, Comparative Physiology of the Vertebrate Digestive System, 2d ed., Cambridge University Press, 1995*)

enzymes that aid in the final digestion of food and mechanisms that provide for the selective absorption of nutrients. The lumenal surface area of the human small intestine is increased 10-fold by the presence of villi and an additional 20-fold by the microvilli, resulting in a total surface area of 310,000 in.^{2} (2,000,000 cm^2), which is larger than a doubles tennis court.

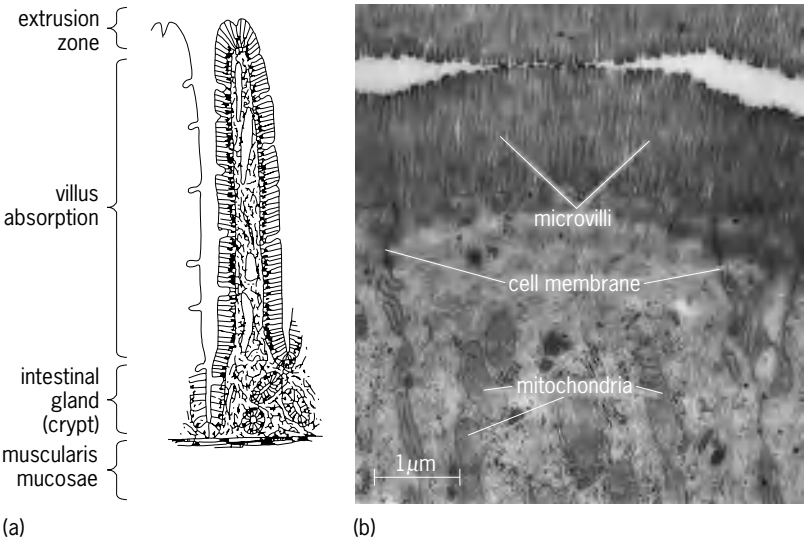

Fig. 3. Section of the wall of the small intestine. (*a*) Intestinal villus and crypt (*after R. A. Argenzio et al., Villous atrophy, crypt hyperplasia, cellular infiltration, and impaired glucose-Na absorption in enteric cryptosporidiosis in pigs, Gastroenterology, 98:1129–1140, 1990*). (*b*) Electron micrograph of microvilli on the free surface of an absorbing cell from the small intestine of a rat (*courtesy of M. L. Watson*).

Pancreas and biliary system. Digestion in the midgut is aided by secretions of digestive enzymes and fluid by pancreatic tissue, and secretion of bile by the liver. Pancreatic tissue is represented by primitive ceca along the midgut of cyclostomes and is distributed along the intestinal wall, and even into the liver, of some species in the more advanced classes of fish. However, the pancreas is a compact organ in sharks, skates, rays, many teleosts, and all other vertebrates. *See* PANCREAS.

The liver is a compact organ in all vertebrates. One of its many functions is the secretion of bile. In most vertebrates, the bile is stored in the gallbladder and released into the intestine as needed, but a gallbladder is absent in some species of fish and mammals. Bile salts serve to emulsify lipids and increase their surface area available for digestion by the water-soluble lipase. *See* GALLBLADDER; LIVER.

Hindgut. The hindgut is the final site of digestion and absorption prior to defecation or evacuation of waste products. The hindgut of fish, amphibian larvae, and a few mammals is short and difficult to distinguish from the midgut by either its diameter or epithelial lining. However, the hindgut of adult amphibians and reptiles, birds, and most mammals is a distinct segment, which is separated from the midgut by a muscular sphincter or valve. It also tends to be larger in diameter. Thus, the midgut and hindgut of these animals are often referred to as the small intestine and the large intestine. *See* INTESTINE.

The hindgut of some reptiles and many mammals includes a blind sac or cecum near its junction with the midgut. A pair of ceca are present in the hindgut of many birds and a few mammalian species. The remainder of the hindgut consists of the colon and a short, straight, terminal segment, which is called the rectum in mammals. The digestive and urinary tracts exit separately from the body of most species of fish and mammals. However, in adult amphibians and the reptiles, birds, and some mammals, this segment terminates in a chamber called the cloaca, which also serves as an exit for the urinary and reproductive systems. The hindgut or, where present, the cloaca terminates in the anus. *See* COLON; URINARY SYSTEM.

The hindgut is similarly lined with a single layer of absorptive and mucus-secreting cells that originate in crypts. However, it lacks villi, and (with the exception of the cecum of birds) its absorptive cells lack digestive enzymes and the ability to absorb most nutrients. One major function of the hindgut is to reabsorb the fluids secreted into the upper digestive tract and (in animals that have a cloaca) excreted in the urine. It also serves as the principal site for the microbial production of nutrients in the herbivorous reptiles and birds and in most herbivorous mammals. Thus, the hindgut tends to be longest in animals that need to conserve water in an arid environment, and has a larger capacity in most herbivores.

Musculature. The digestion of food, absorption of nutrients, and excretion of waste products require

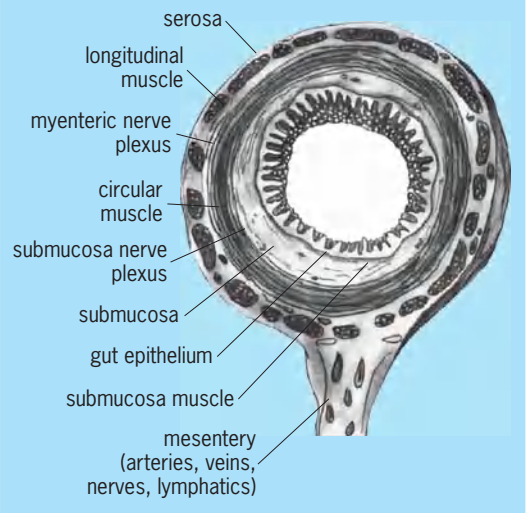

Fig. 4. Cross section of the layers of the gastrointestinal tract and the mesentery, which supplies it with blood vessels, lymphatic vessels, and nerves.

the mixing of ingesta with digestive enzymes and the transit of ingesta and digesta through the digestive tract. The digestive tract of cyclostomes is enveloped with a single layer of muscle, and digesta transit is aided by cilia in the epithelial lining of the tract. However, in other vertebrates the contents are mixed and moved by an inner layer of circular muscle and an outer layer of muscle that runs longitudinally along the tract (**Fig. 4**).

The initial act of deglutition and the final act by which waste products are defecated from the digestive tract are effected by striated muscle. This type of muscle is characterized by rapid contraction and is controlled by extrinsic nerves. The esophagus of fish and varying lengths of the mammalian esophagus are enveloped with striated muscle. However, the esophagus of amphibians, reptiles, and birds, and the entire gastrointestinal tract of all vertebrates are enveloped by smooth muscle. This smooth muscle contracts more slowly, and its rate of contraction is partly independent of external stimulation. *See* MUSCLE.

Contraction of the circular muscles aids in the mixing and propulsion of digesta. At some sites along the digestive tract, the circular muscle is thicker and under almost continuous contraction, forming sphincters or valves that delay the movement of lumen contents in either direction. The anus has an additional sphincter, which is composed of striated muscle.

Contractions of the longitudinal muscles shorten the digestive tract. In the large intestine of some

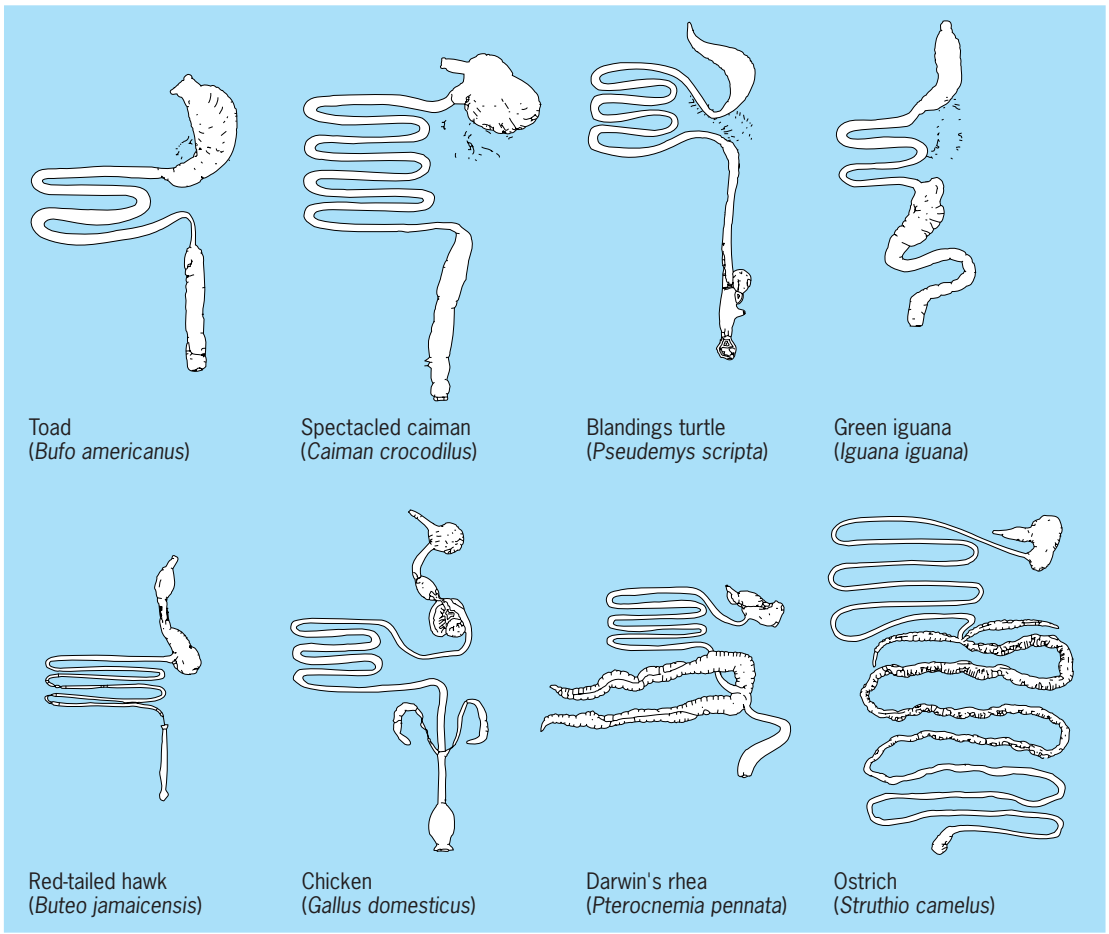

Fig. 5. Gastrointestinal tracts of amphibians, reptiles, and birds. (*After C. E. Stevens and I. D. Hume, Comparative Physiology of the Vertebrate Digestive System, 2d ed., Cambridge University Press, 1995*)

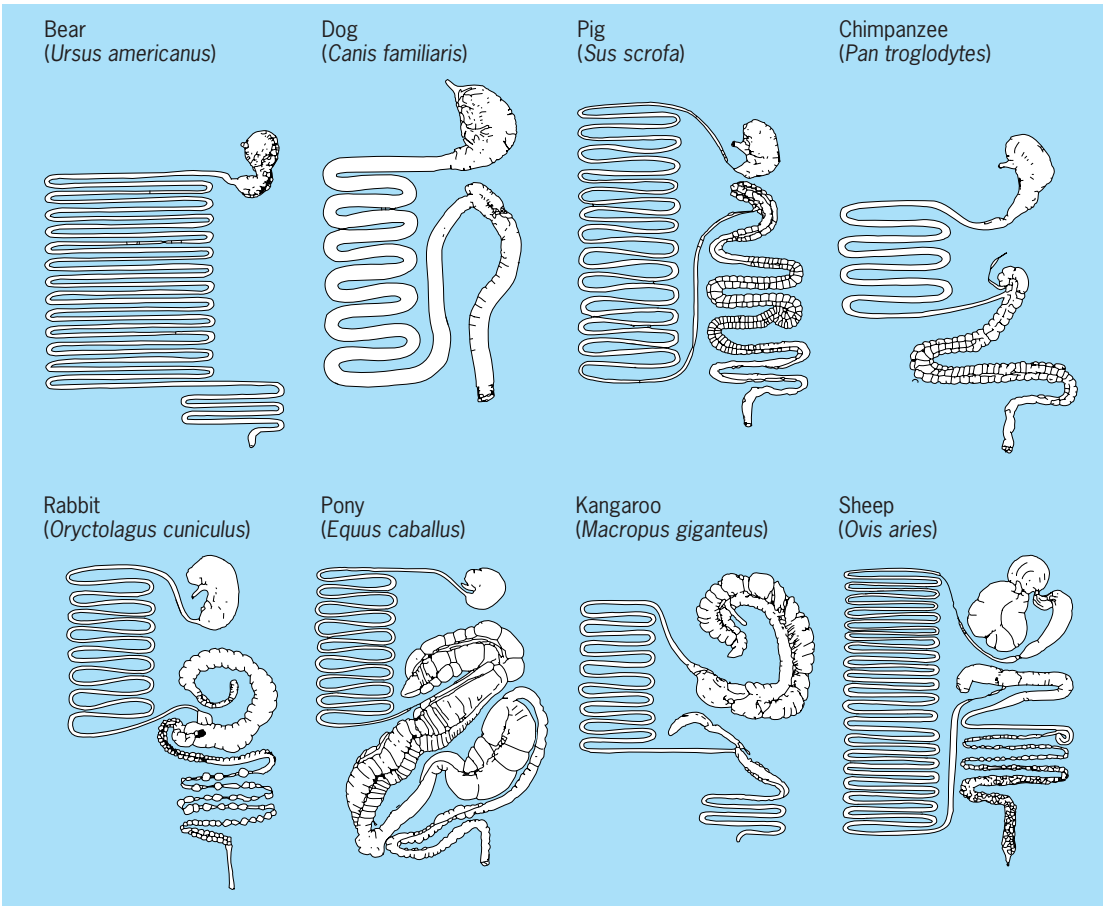

Fig. 6. Gastrointestinal tracts of carnivorous, omnivorous, and herbivorous mammals. (*Bear and chimpanzee after C. E. Stevens and I. D. Hume, Comparative Physiology of the Vertebrate Digestive System, 2d ed., Cambridge University Press, 1995; pig after R. A. Argenzio and M. Southworth, Sites of organic acid production and absorption in the gastrointestinal tract of the pig, Amer. J. Physiol., 118:454–460, 1974; dog, rabbit, pony, kangaroo, and sheep after C. E. Stevens, Comparative physiology of the digestive system, in Dukes' Physiology of Domestic Animals, ed. by M. J. Swenson, 9th ed., pp. 216–232, Cornell University Press*)

mammals and the stomach of a few, the longitudinal muscle is concentrated into three bands (taenia). Contraction of these longitudinal bands pulls these segments of the gut into a series of saccules. These saccules change their position along the gut with time and serve to both mix the contents and propel them through the gut.

Nerve and endocrine tissue. The initial act of deglutition and final act of defecation are under the voluntary control of the central nervous system. However, the remainder of the digestive system is subject to the involuntary control of nerves which release a variety of neurotransmitting or neuromodulating agents that either stimulate or inhibit muscular contractions and the secretions of glands and cells. The motor and secretory activities of the digestive system are also under the control of a wide range of other substances produced by endocrine cells that are released either distant from (hormones) or adjacent to (paracrine agents) their site of action. Although there are some major variations in the complement and activities of the neurotransmitters, neuromodulators, hormones, and paracrine agents, their basic patterns of control are similar.

Variations. The gastrointestinal tracts of adult amphibians and reptiles consist of a simple stomach, a midgut, and a hindgut that is generally expanded in diameter and referred to as the large intestine (**Fig. 5**). The hindgut is relatively short in adult amphibians, such as the toad, and in reptilian carnivores, such as the caiman. It tends to be longer in omnivorous reptiles, such as Blanding's turtle, and the proximal hindgut of most herbivorous reptiles, such as the green iguana, is expanded and includes a cecum at its juncture with the midgut. The gastrointestinal tract of most birds includes a crop, proventriculus, gizzard, paired ceca, and a short straight section of hindgut that is generally referred to as the rectum. However, ceca are absent in some birds, small in others, such as the hawk, and more developed in omnivores, such as chickens. The largest ceca are found in most of the herbivores, such as grouse and rheas. The long colon of the herbivorous ostrich is unique among avian species.

Mammals. The stomach and hindgut of mammals show an even wider range of structural variation (**Fig. 6**). Most mammals have a simple stomach, but the forestomach of many herbivorous mammals is large and divided into numerous compartments or

saccules, as seen in the sheep and kangaroo. The same is true for other advanced ruminants (bovids, goats, antelope, deer, elk, giraffe), camelids, hippopotamuses, as well as the sloths, colobid monkeys, and cetaceans (dolphins, porpoise, and whales), hyrax, and some herbivorous rodents. Although present-day cetaceans feed on invertebrates, fish, or small mammals, the multicompartmental stomach is believed to have been inherited from herbivorous ancestors. The stomach is also expanded in blood-feeding vampire bats, reaching twice the body length of some species.

The mammalian hindgut shows a similar range of variation. It is very short and indistinct in some insectivores, cetaceans, and the omnivorous bears, and relatively simple in most carnivores, such as the dog. The hindgut of omnivores tends to be more highly developed, and includes a sacculated cecum and colon in some omnivores, such as the chimpanzee and domesticated pig. The hindgut of mammalian herbivores with a simple stomach is dominated by either a large, sacculated cecum, as seen in the rabbit, or a large sacculated colon, such as that of the pony. However, regardless of diet, the hindgut tends to be short and simple in aquatic and semiaquatic species, such as cetaceans and hippos, and longest in desert species, such as camels.

Human. The anatomy of the human digestive system is similar to that of other mammalian omnivores (**Fig. 7**). The teeth and salivary glands are those of a mammalian omnivore, and the initial two-thirds of the esophagus is enveloped by striated muscle. A simple stomach is followed by an intestine, whose length consists of approximately two-thirds small bowel and one-third large bowel. The structures of the pancreas and biliary system show no major differences from those of other mammals. During early fetal development, a distinct, conical cecum is present and continues to grow until the sixth month of gestation. However, unlike other primates, the cecum recedes to become little more than a bulge in the proximal colon by the time of birth. The colon continues to lengthen after the birth and is sacculated throughout its length like that of the apes and a few monkeys but few other mammals.

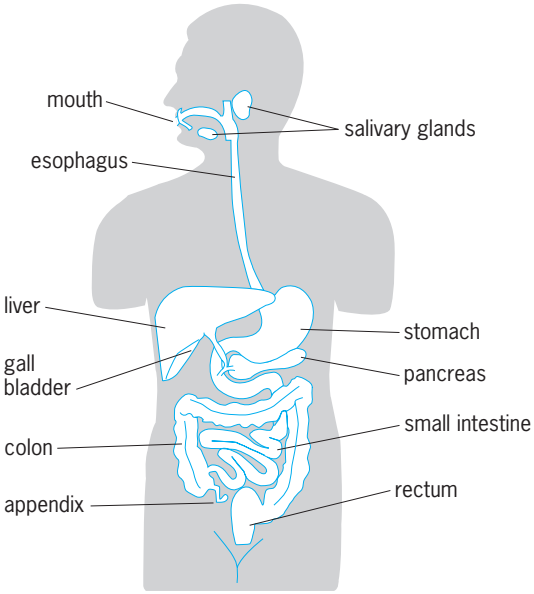

Fig. 7. Human digestive system. (*After A. C. Guyton, Textbook of Medical Physiology, 8th ed., Saunders, 1991*)

Physiology

The major physiological activities of the digestive system are motility, secretion, digestion, and absorption. Each activity can be affected by diet and, in the cold-blooded species, is reduced with a decrease in body temperature.

Motility. The mastication of food and the movement of ingesta and digesta through the digestive tract are controlled by the motor activity of muscular contractions. Pressure of food against the palate and back of the mouth stimulates a nerve reflex that passes through a deglutition center in the brain. This reflex closes the entrance into the respiratory system and stops respiration, to prevent the inspiration of food into the lungs, and initiates muscular contractions that pass food into the esophagus. The food (bolus) is then passed down the esophagus and into the stomach by a moving wave of muscular contractions (peristalsis) accompanied by inhibition of the esophageal sphincters.

The multicompartmental forestomach of ruminants undergoes a continuous series of complex, repetitive contractions that are controlled by the central nervous system. However, the gastric motility of most species and the intestinal motility of all vertebrates are controlled partially by the intrinsic characteristics of their smooth muscle cells. The result is production of either stationary (mixing) contractions of the stomach and intestine or a series of peristaltic contractions that carry digesta on through the tract.

Gastric peristalsis is initiated by cells in the middle of the stomach, which act as pacemakers for the contractions that travel over the distal half of the stomach. Pacemakers in the midgut and hindgut can also initiate waves of peristaltic contraction. However, pacemakers located near the cloaca of reptiles and birds or the midcolon of most mammals generate periodic waves of antiperistaltic contraction, which reflux digesta along the colon and into the cecum. These antiperistaltic waves increase the retention time needed for the absorption of electrolytes and water, and the multiplication of hindgut bacteria. Peristaltic and segmental contractions eventually move the digesta into the terminal hindgut, where it initiates the relaxation of sphincters and defecation.

Digestion. The diet of vertebrates contains carbohydrates, lipids, and protein, which provide the energy and building blocks for growth, maintenance, and reproduction, and the vitamins that are required in small amounts for the regulation of metabolic processes. The lipids, proteins, and most of the carbohydrates require digestion for the release of nutrients that can be absorbed. Digestion is accomplished by enzymes produced by the digestive

Major dietary substrate end products and endogenous digestive enzymes*

Substrate	Extracellular enzymes	Intestinal mucosal enzymes	End products
Amylose Amylopectin Glycogen	α Amylase	Maltase Isomaltase	Glucose
Chitin	Chitinase	Chitobiase	Glucosamine
Sucrose		Sucrase	Glucosamine
Lactose		Lactase	Glucose Galactose
Trehalose		Trehalase	Glucose
Triglycerides	Lipase Colipase		β-Monoglyceride Fatty acids
Waxes	Lipase Esterase		Monohydric alcohol Fatty acids
Protein	Pepsin Trypsin Chymotrypsin Elastases Carboxypeptidase A Carboxypeptidase B	Aminopeptidase Tripeptidase Dipeptidase	Amino acids

*Modified from C. E. Stevens and I. D. Hume, *Comparative Physiology of the Vertebrate Digestive System,* 2d ed., Cambridge University Press, 1995.

system (endogenous enzymes) or by bacteria that are normal residents of the digestive tract.

Dietary substrates. The carbohydrates consist of monosaccharides, disaccharides, and polysaccharides. The major monosaccharides in plants and animals are glucose or fructose, and the major disaccharides are the transport carbohydrates: sucrose (glucose-fructose), lactose (glucose-galactose), and trehalose (glucose-glucose). Sucrose is the principal transport carbohydrate in plants, trehalose is the blood sugar of insects, and lactose is the milk sugar of most mammals. Polysaccharides serve as either storage or structural carbohydrates. The main storage carbohydrates are the starches amylose and amylopectin in plants and glycogen in animals. The principal structural carbohydrates are the cellulose, hemicellulose, and pectin in the cell walls of plants, and chitin, which reinforces the exoskeleton cell walls of insects and many marine invertebrates. Starch and cellulose consist of long chains of glucose molecules. Hemicellulose, pectin, and chitin are made up of a variety of monosaccharides.

The major dietary lipids (fats and oils) are triglycerides and waxes that consist of esters of alcohols and fatty acids. Triglycerides are the major form of lipid storage in most animals and the seeds of plants. Wax esters constitute most of the lipids stored in many of the invertebrates in the food chain of marine fish and mammals. Protein consists of long chains of amino acids, arranged in various combinations and permutations that must be broken down to small peptides or amino acids for intestinal absorption. The cells of plants, animals, and bacteria also contain deoxyribonucleic acid (DNA), which provides genetic information, and ribonucleic acid (RNA), which controls the activities of the cells. The vitamins required by vertebrates consist of a group of fat-soluble (A, D, E, and K) and water-soluble (B-complex and C) organic compounds.

Endogenous enzymes. Plant and animal starches are converted to oligosaccharides (short-chain structures) and disaccharides by amylase, which is secreted by the salivary glands of some species and the pancreas of all vertebrates (see **table**). The end products of starch digestion, plus the dietary disaccharides, are converted to monosaccharides by enzymes in the brush border of the absorptive epithelial cells lining the small intestine. Chitin can be digested by the enzyme chitinase, which has been demonstrated in the gastric mucosa and pancreas of many mammals, birds, reptiles, and adult amphibians. However, vertebrates do not produce enzymes capable of digesting the structural polysaccharides of plants.

Lipids are digested into alcohols, monoglycerides, and fatty acids by lipases and esterases, which are secreted predominantly by the pancreas. However, the lipases are water-soluble enzymes that can attack their substrate only at a lipid-water interface. Therefore, the lipids must be emulsified in order to provide the surface area required for efficient digestion. Emulsification is accomplished by the release of bile salts secreted by the liver and released into the midgut. The bile salts help form micelles, small spherical structures that contain the end products of lipid digestion and aid in their transfer to the lumen membrane of the epithelial cells for absorption into the cell. Once this process is complete, the bile salts are released and passed on to the distal midgut, where they are absorbed and circulated back to the liver.

Dietary protein is first broken down into long chains of amino acids (polypeptides) by gastric pepsin and pancreatic trypsin. The polypeptides are then attacked by other pancreatic proteases (chymotrypsin, carboxypeptidase, elastase) to form tripeptides, dipeptides, and amino acids. All of these enzymes are secreted in an inactive form to prevent the self-digestion of the secretory cells prior to their release. Pepsin is activated by the acidity resulting from the secretion of hydrochloric acid (HCl) into the stomach, and trypsin is activated by an enzyme (enterokinase) that is secreted by intestinal epithelium. The trypsin initially released by enterokinase activates the release of additional trypsin and the other pancreatic, proteolytic enzymes. Tri- and dipeptides are digested into amino acids by enzymes in the brush border and contents of midgut absorptive cells. Nucleic acids are digested by pancreatic ribonucleases into pentose sugars, purines, and pyrimidines.

Some vitamins must be released from their combination with other substances by digestive enzymes before they can be absorbed. Vitamin B_{12}, which is especially subject to destruction by digestive enzymes, is released from other compounds and combined with a substance called intrinsic factor secreted by the stomach. This combination protects it from digestion until it is released for absorption in the distal midgut. Therefore, a B_{12} deficiency can result from dysfunctions of gastric secretion in humans.

The levels of various digestive enzymes tend to correlate with the levels of carbohydrate, lipid, and protein in the normal diet of a species. Herbivores and omnivores tend to produce higher levels of amylase, carnivores tend to produce higher levels of proteases, and the high concentrations of ribonucleases found in ruminants and kangaroos correlate with the high levels of microbial nucleic acid produced in their forestomach. The composition of digestive enzymes also varies with time in neonate mammals. All neonate mammals feed on milk, which is high in lipids and lactose in most species, and after weaning their diet that is usually low in fat and higher in other carbohydrates. The production of digestive enzymes adjusts to these changes in the diet. For example, the newborn of most species produce high levels of lactase but low levels of the other carbohydrases. However, the latter appear well before weaning and reach their full complement at the time of birth, with a subsequent decrease in lactase production. These changes in enzyme composition anticipate the changes in diet and appear to be genetically programmed or "hard-wired" into the system. The levels of pancreatic amylase, lipase, and proteases (brush border enzymes) can also increase following an increase in the carbohydrate, lipid, or protein in the diet of adult animals.

Indigenous bacteria. Soon after birth or hatching the lumen surface and contents of the gastrointestinal tract become colonized by bacteria. Substantial numbers of bacteria can be found in all segments of the tract, but the highest numbers are present in those segments in which digesta are retained for prolonged periods of time at a relatively neutral pH. Bacteria are found in concentrations of 10^4 to 10^6 microorganisms per gram of fluid digesta in the midgut of fish and higher vertebrates. However, the development of the hindgut of terrestrial vertebrates into a major site of digesta retention was accompanied by its colonization with much higher numbers (10^{10} to 10^{11} per gram). In the ruminants and a few other species of mammalian herbivore, a large stomach serves as an additional site for colonization with similar numbers of bacteria.

Indigenous bacteria help protect the animal from pathogenic microorganisms by stimulating immunity and competing for substrates. They also convert dietary and endogenous substances that are not digested by endogenous enzymes into absorbable nutrients. Many species of indigenous bacteria can ferment sugars, starches, and structural carbohydrates into short-chain fatty acids. The short-chain fatty acids, which are predominantly acetic, propionic, and butyric acids, are readily absorbed and serve as an additional source of energy. These bacteria also synthesize microbial protein and the B-complex vitamins that may be useful to their host. *See* BACTERIAL PHYSIOLOGY AND METABOLISM.

The nutritional contributions of hindgut bacteria to mammals are illustrated in **Fig. 8**. Hindgut bacteria produce short-chain fatty acids by the fermentation of dietary carbohydrates that have escaped digestion and absorption in the midgut, and endogenous carbohydrate, such as that present in mucus. They also utilize dietary and endogenous compounds, such as digestive enzymes and urea, for the production of ammonia and microbial protein. Although most of the urea, which is the waste product of protein metabolism by the mammalian liver, is excreted by the kidneys, substantial amounts enter the gut by diffusion across its epithelial lining. The hindgut bacteria of birds, reptiles, and adult amphibians perform similar functions, except that uric acid is the waste product of hepatic protein metabolism and it enters the hindgut via the cloaca rather than by diffusion. The bacteria in the forestomach of mammalian herbivores also produce short-chain fatty acids from dietary carbohydrates, and ammonia and microbial protein from dietary protein and from urea, which reaches the forestomach via the saliva and diffusion across its epithelial lining.

Absorption of short-chain fatty acids provides the major source of the energy required by hindgut and forestomach epithelial cells and a portion of the maintenance energy required by the species. Some of the ammonia is used for the synthesis of bacterial protein, but most of the remainder is absorbed and returned to the liver for the synthesis of protein and urea. This aids in the conservation of nitrogen and reduces the amount of urea (and water) that must be excreted in the urine. The absorption of short-chain fatty acids and ammonia also plays a critical role in the hindgut absorption of water. The microbial protein and B-complex vitamins synthesized by these bacteria also can be utilized as nutrients by coprophagic and forestomach fermenting species.

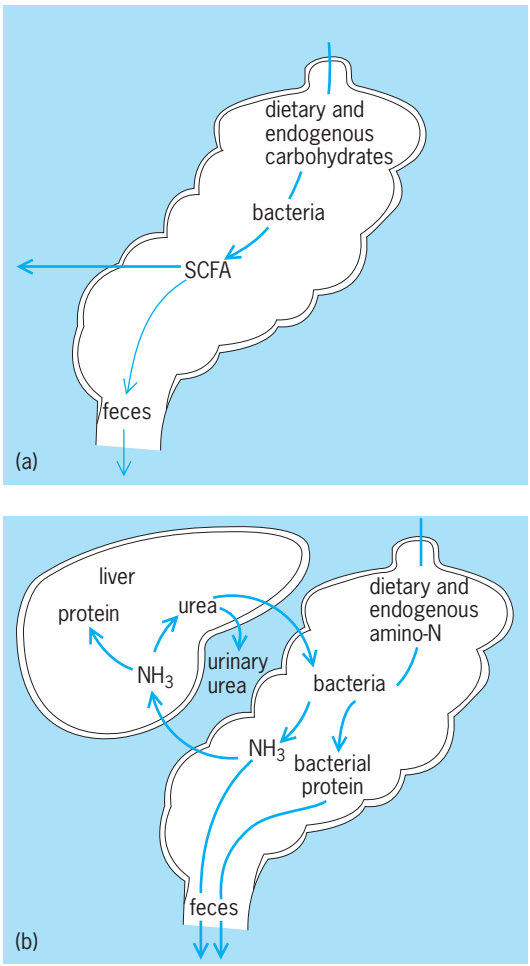

Fig. 8. Bacterial fermentation of carbohydrates and metabolism of nitrogenous compounds in the hindgut of mammals. (*a*) Fermentation of dietary and endogenous carbohydrates to short-chain fatty acids (SCFA), which are largely absorbed. (*b*) Bacterial production of ammonia and microbial protein from dietary and endogenous nitrogenous compounds. The endogenous compounds include digestive enzymes released in the upper digestive tract and the urea that passively diffuses into the gut. Ammonia is absorbed and returned to the liver for incorporation into protein and urea, which serves to conserve nitrogen and reduce the amounts of urea and water excreted by the kidneys. Bacteria in the hindgut of reptiles and birds serve similar functions, except that the uric acid that is produced instead of urea enters the hindgut via the cloaca. (*Modified from A. M. Wrong and A. J. Vince, Urea and ammonia metabolism in the human large intestine, Proc. Nutrit. Soc., 43:77–86, 1984*)

The contributions of indigenous bacteria to the production and conservation of nutrients are greatest in herbivores. Although it has been estimated that short-chain fatty acid absorption provides 4% of total maintenance energy requirement by dogs and 6–10% of the maintenance energy required by humans, they can account for 30% of the maintenance energy of rabbits and up to 70% of maintenance energy of horses and ruminants. The dietary protein and water requirements of rabbits, horses, and ruminants can be greatly reduced by the recycling of urea nitrogen; and the dietary requirements for both protein and B-vitamins are substantially reduced by coprophagy in rabbits and foregut fermentation in ruminants.

Herbivores. The appearance of herbivores that could subsist largely on the leaves, petioles, and stems of plants played a significant role in the expansion and distribution of vertebrates. Herbivory requires the ingestion of large quantities of plant material, its reduction to small particle size, and either the rapid passage of this material through the gut or its prolonged storage for bacterial fermentation. Herbivores are relatively rare among fish, larval amphibians, reptiles, and birds, and all adult amphibians are carnivores. Most herbivorous fish reduce plant material to particle size by the use of a pharyngeal mill or gizzardlike stomach and pass it rapidly through the intestine, using their midgut as the principal site of microbial fermentation. The midgut is also the principal site in larval amphibians and the Florida red-bellied turtle, emu, and panda, but the colon is the principal site in most herbivorous reptiles.

The rarity of herbivorous reptiles has been attributed to the lack of an efficient masticatory apparatus and small body size of most reptiles. The articulation of the reptilian jaws precludes the grinding action that can reduce plant material to a fine size, and the increase in mass-specific metabolic rate with a decrease in body mass limits the gut capacity and, thus, minimum body size of most reptilian herbivores to about 300 g (11 oz). This rarity of herbivores among present-day reptiles has raised questions about why herbivory was so successful for dinosaurs. The sauropods included many herbivores and the largest species of dinosaurs, and all ornithopods are believed to have been herbivores. Their success has been attributed partly to a gizzardlike stomach in the sauropods and an efficient masticatory apparatus in the ornithopods. It has also been attributed to the possibility that many dinosaurs were endothermic. However, studies of present-day reptiles suggest that the lower rates of metabolism and digesta passage that accompany a reduction in the ambient temperature may compensate for the reduction in the rates of digestion and absorption. Therefore, the rarity of present-day herbivores may be due to lack of an efficient masticatory apparatus and the susceptibility of larger species to predation by endotherms.

Although the gizzard of birds provides an efficient apparatus for the grinding of plant material, their gut capacity is limited by the requirement for flight. The herbivorous birds are confined to grouse, partridge, and the hoatzin, which fly only short distances, and the flightless rheas, emus, and ostriches. The midgut is the main site of microbial fermentation in the emu, and the colon is the principal site in the ostrich. However, among the smaller herbivores, the ceca are the principal sites in the rhea, grouse, and partridge, and the foregut (crop and distal esophagus) is the principal site in the hoatzin. The cecal fermenters include snow partridge and snow cocks, which inhabit altitudes 5000 m (16,400 ft) above sea level, and the arctic ptarmigan.

In contrast to the fish, amphibians, reptiles, and birds, herbivory was adopted by a high percentage of mammals. Herbivorous species are found in 11

of the 20 mammalian orders, and 5 of these orders contain only herbivores. The herbivorous mammals constitute the largest number of species, include the largest terrestrial species, show the widest range of ecological distribution, and provide over 90% of the mammalian biomass. Their success is due to an efficient masticatory apparatus and the development of a hindgut or foregut that can store large amounts of plant material for long periods of time.

The principal site of microbial fermentation in the herbivorous panda is a long midgut, similar to that of the bear (Fig. 6), but other mammalian herbivores have adopted the strategies of colon, cecal, or foregut fermentation. Many herbivores over 30 kg (66 lb) in body weight, including the wombats, orangutans, gorillas, tapirs, equids, rhinos, and elephants, have adopted the colon as principal site of microbial fermentation. However, most herbivores less than 30 kg in body weight are cecal fermenters, and most herbivores of intermediate body weight are foregut fermenters.

As with the birds, most small mammalian herbivores overcame the limitations of body weight on gut capacity by the rapid passage of large digesta particles through the gut and selective retention of fluid, bacteria, and small particles in an enlarged cecum. Included are the lagomorphs (rabbits, hares, and pica), folivorous marsupials (koalas, greater gliders, cuscuses, and some possums), and the herbivorous rodents. Many of these species practice coprophagy, which recovers microbial protein and B-vitamins from the feces; and some of these species ingest only the nutritionally rich feces that are derived from a periodic release of cecal contents. Lagomorphs and herbivorous rodents are distributed throughout all regions of the world. The forestomach is the principal site of microbial fermentation in a few small herbivores (rat kangaroos, sloths, colobid monkeys, and the smallest ruminants) and most intermediate-sized herbivores (kangaroos, wallabies, camelids, and advanced ruminants).

Colon and cecal fermenters have the advantages of recovering nutrients from the readily digestible components of the diet from their midgut, prior to the less efficient process of microbial digestion, and increasing their food intake, digesta passage, and recovery of these nutrients on high-fiber diets. However, foregut and coprophagic cecal fermenters make use of the protein and B vitamins synthesized by gut microbes. Foregut fermentation sacrifices the advantage of recovering readily available nutrients prior to microbial fermentation. However, it allows the microbial digestion of plant toxins and provides the most efficient system for the recovery of large amounts of the electrolytes and water secreted by herbivores.

Ruminants have the additional ability of selectively retaining large particles of plant fiber in their forestomach. This retention allows a more complete fermentation of low-quality forage and accounts for the ruminants' adaptation to desert, high altitudes, and the arctic tundra. Therefore, the adoption of cecal and foregut fermentation expanded herbivory into smaller birds and mammals, and extended their distribution into regions where forage and water are sparse.

Absorption. The cells lining the small intestine of many newborn mammals are capable of absorbing intestinal contents by a process called phagocytosis. Intestinal contents are engulfed by the lumen-facing membrane into cell vesicles. Cellular enzymes digest most of the vesicular contents, but the immunoglobulins in the mother's milk are transported intact and provide the newborn with a passive immunity to many diseases. Phagocytosis is absent in the newborn of some species and short-termed in others, such as humans. Although it can continue for longer periods in a few species, such as the rat, it disappears in all mammals by the time of weaning.

The epithelial cells that line the gastrointestinal tract are closely attached to one another at their lumen-facing border by tight junctions, which are relatively impermeable to most substances other than water. Therefore, the major restriction for the absorption of most substances from the lumen into the blood is the apical and basolateral membranes of these cells. The epithelial cell membranes consist of a lipid-protein barrier that is permeated with a limited number of small, water-filled pores. Thus, they are relatively permeable to lipid-soluble substances and water but relatively impermeable to the passive diffusion of most water-soluble substances. *See* CELL PERMEABILITY.

Lipid-soluble substances. Lipid-soluble substances can be transported across the apical cell membranes by passive diffusion down their concentration gradient. The short- and medium-chain fatty acids that result from lipid digestion in the small intestine pass directly into the blood. However, the monoglycerides and long-chain fatty acids are resynthesized into triglycerides by the epithelial cells in the midgut and incorporated into small spheres (chylomicrons), which are transported across the basolateral membrane into the lymphatic system. Fat-soluble vitamins, long-chain alcohols, and other lipids also appear to be incorporated into chylomicrons and to enter the lymphatic system.

Water-soluble compounds. The intestinal cell membranes are relatively impermeable to the passive diffusion of water-soluble monosaccharides, amino acids, vitamins, and minerals that constitute a major portion of the required nutrients. These nutrients are selectively transferred across the intestinal cell membranes by carrier-mediated transport. Membrane carriers combine with the nutrient at one membrane surface and pass it across the membrane for release at the opposing surface. Some simply facilitate the diffusion of a substance down its concentration gradient; others are capable of transporting a nutrient against its concentration gradient, which requires either a direct or indirect investment of cellular energy. *See* CELL MEMBRANES.

Fructose is transported by a carrier-mediated process that facilitates its diffusion across the apical membrane, but glucose and galactose are transported into the epithelial cells of the small intestine

against their respective concentration gradients. Their absorption into the cell is dependent upon the simultaneous absorption of sodium ions (Na^+) down the concentration gradient provided by transport of sodium by another carrier against its concentration gradient across the basolateral membrane. This relationship between glucose and sodium absorption was first suggested by the finding that cola beverages, which contain high levels of glucose and sodium, were much more effective than glucose solutions when administered orally for treatment of the diarrhea associated with cholera. The amino acids, dipeptides, and B-complex vitamins also are transported across the apical membrane of these intestinal cells by a variety of Na^+-dependent, carrier-mediated transport systems. The transport of these substances across the basolateral membranes of the cell is accomplished by sodium-independent carriers.

Short-chain fatty acids and ammonia. The short-chain fatty acids and ammonia derived from bacterial fermentation are present in both undissociated and dissociated (ionized) form. Cell membranes are relatively permeable to the passive diffusion of the lipid soluble undissociated short-chain fatty acids and ammonia, but impermeable to the diffusion of the water-soluble ionized forms that predominate at the normal pH of hindgut and forestomach contents. Therefore, a reduction in digesta pH tends to increase the rate of short-chain fatty acid absorption and decrease the rate of ammonia absorption. However, there is also evidence for the absorption of short-chain fatty acid ions in exchange for bicarbonate (HCO_3^-) and ammonia ions (NH_4^+) in exchange for hydrogen (H^+).

Minerals. The metabolic processes of the body require a number of different minerals. Some such as iron, calcium, sodium, and chloride are required in relatively large quantities. Others such as manganese and zinc are labeled trace minerals because they are required in only minute amounts. The absorption of iron and calcium involves membrane carriers. Absorption of sodium and chloride ions is discussed below.

Diet adaptations. The rates of nutrient absorption can be affected by the diet. The ratio of sugar to amino acid absorption appears to be highest in the herbivores and lowest in the carnivores of all classes of vertebrates. A change from a high-carbohydrate and low-protein diet to a high-protein and low-carbohydrate diet also can increase the rate of glucose absorption and decrease the rate of amino acid absorption by individual animals. An increase in dietary calcium, iron, copper, phosphorus, or B-complex vitamins, which can be toxic at high concentrations, appears to decrease their rates of absorption.

Electrolytes and water. The nutrient that is required in largest quantity for digestion, absorption, metabolism, and excretion of waste products is water. Because it readily diffuses across cell membranes down its concentration gradient, the net secretion or absorption of water is determined by the net secretion or absorption of all other substances. Sodium, chloride, and bicarbonate are the principal ions that are present in the extracellular fluids that bathe the body cells of all vertebrates and that are transported across cell membranes. Therefore, the transport of these electrolytes is the major driving force for the secretion or absorption of water.

Secretion. Solubilization, digestion, and the transit of food and digesta by the alimentary tract require the secretion of large volumes of fluid. Secretions by the salivary glands, pancreas, biliary system, and intestinal cells contain high concentrations of sodium, chloride, and bicarbonate ions, and gastric secretions contain high levels of hydrogen and chloride ions. The secretion of hydrochloric acid reduces the pH of gastric contents to the levels required for activation and activity of pepsin. However, the midgut requires a more alkaline pH for protection of its epithelium and the activation of the pancreatic enzymes. This is accomplished by the bicarbonate ions secreted by the pancreas, biliary system, and intestinal mucosa. The short-chain fatty acids produced by microbial fermentation can also reduce the pH of gut contents, with damaging results. Short-chain fatty acids produced in the forestomach are neutralized by bicarbonate secreted from the salivary glands and the forestomach epithelium, and those produced in the hindgut are neutralized by the bicarbonate secreted into the midgut and hindgut. *See* PH REGULATION (BIOLOGY).

Under normal conditions, an adult human consumes 2 liters (1.8 qt) of water per day, and the digestive system secretes an additional 7 liters (6.4 qt) of fluid per day; resulting in a total of 9 liters (8.2 qt) of fluid. This is equivalent to 40% of the water in the extracellular fluid space. An uncompensated loss of 15% of the extracellular fluid volume can produce clinical signs of dehydration, and a loss of 35% can result in death. This is prevented by absorption of 7 liters of this fluid by the small intestine and an additional 1.4 liters (1.3 qt) by the large intestine, leaving only 0.1 liter (0.09 qt) for excretion into the feces. Although the large intestine normally absorbs only 20% of this fluid, this percentage is critical, and its absorptive capacity can be doubled if necessary. The daily secretions of the digestive system are much higher in herbivores, reaching levels equivalent to two or three times the extracellular fluid in the sheep and horse, but with a similar 95–98% recovery prior to excretion.

Absorption. The absorption of water from the midgut is accomplished mainly by the absorption of sodium, chloride, vitamins, and the end products of carbohydrate, lipid, and protein digestion by the endogenous enzymes. Some of the sodium is absorbed in conjunction with monosaccharide, amino acid, and vitamin absorption, but some of the sodium and most of the chloride are absorbed by a mechanism that exchanges sodium for the hydrogen and chloride for the bicarbonate in the epithelial cells. The hydrogen and bicarbonate entering the gut combine to form carbon dioxide (CO_2) and water.

The hindgut appears to have evolved initially for conservation of the electrolytes, water, and nitrogen in both the digestive secretions and urinary excretions of terrestrial vertebrates. The kidney is the

principal organ for excretion of metabolic waste products, which it accomplishes by continuously filtering blood through the glomerulus of the nephron and reabsorbing water and nutrients from the renal tubule. The earliest vertebrates are believed to have been fresh-water fish, whose major problem was the rejection or elimination of excess water. Fresh-water fish do not intentionally drink water, and excess water is excreted by the kidneys.

Marine fish face the opposite problem: the potential loss of water and gain in sodium chloride (NaCl) due to the high salinity and osmotic activity of seawater. Although marine fish drink seawater, its osmotic effect on the digestive system is reduced by the absorption of sodium chloride from their esophagus. Hagfish, sharks, and rays have resolved the osmotic problem by having body fluids that are isotonic or hypertonic to seawater. However, this problem is resolved in most marine fish by the absence of glomerular filtration by the kidney and active secretion of sodium chloride by the gills and other tissues such as the rectal glands of sharks. *See* KIDNEY.

Terrestrial animals must conserve both their water and electrolytes against loss to the environment. Their kidneys, like those of fresh-water fish, filter blood and reabsorb electrolytes and water, but the urine of amphibians, reptiles, and birds is excreted into a cloaca. Antiperistaltic contractions then reflux urine along the length of the hindgut of many reptiles and birds, allowing the cloaca and hindgut to aid the kidney in the recovery of electrolytes and water from both the digestive secretions and urinary excretions.

In most mammals, the digestive and urinary systems exit the body separately, and the nephron of the kidney is lengthened by the loop of Henle, which provides for more efficient recovery of water. The hindgut continues to function as the final site for recovery of digestive fluids, and tends to be longer than that of lower vertebrates and longest in those that inhabit arid environments. Pacemaker tissue located in the midcolon generates antiperistaltic waves of contraction, which continue to aid retention of digesta in the cecum and proximal colon.

The absorption of water from the hindgut is due principally to the absorption of sodium, short-chain fatty acids, and ammonia. Sodium is absorbed in exchange for hydrogen by a mechanism that partly depends on short-chain fatty acid absorption and by a mechanism that transports sodium alone. Thus, prolonged retention of digesta by the hindgut of terrestrial vertebrates increased both the time available for absorption of electrolytes and water, and the multiplication of endogenous bacteria that contribute to this process.

Neural and endocrine control. The motor and secretory activities of the digestive system are controlled and integrated by the nervous system, hormones, and paracrine secretions. Many of these activities are controlled by nerve reflexes, consisting of sensory neurons with receptors in the digestive tract that pass information, via other neurons in the central nervous system and gut wall, back to the digestive tract. The latter neurons release neurotransmitters or modulators that stimulate or inhibit muscular contraction or cellular secretions. Reflexes responsible for the initial stage of deglutition and final stage of defecation are monitored through brain centers, allowing for voluntary control of these activities. However, most of the motor and secretory events in the digestive system are under the involuntary and sometimes interactive control of a large number of neurotransmitters, neuromodulators, hormones, and paracrine agents.

Although much of the neuroendocrine control provides a direct response to conditions in a given segment of the gut, some of these responses prepare a following segment for what is to come, or tune down the activities in a previous segment. The sight or smell of food can stimulate gastric secretions, and the act of deglutition produces a receptive relaxation of the stomach. Distention of the stomach can stimulate contractions of the colon, which accounts for the relationship between meals and defecation. Similarly, the entrance of partially digested food into the small intestine not only stimulates the secretion of pancreatic fluid and the release of bile but also decreases the rates of gastric secretion and motility.

For example, the sight or smell of food and its presence in the stomach result in neural and chemical stimulation of the release of the hormone gastrin from the stomach wall. Gastrin stimulates the secretion of hydrochloric acid and pepsinogen for the digestion of protein. As the digestive process progresses, the passage of hydrochloric acid and the end products of gastric digestion into the small intestine stimulates the release of the hormones secretin and cholecystokinin from the intestinal wall. These hormones act to inhibit the secretion of gastrin and stimulate the release of fluid and enzymes from the pancreas and the release of bile from the gallbladder. The high concentration of bicarbonate ions in these secretions neutralizes the hydrochloric acid in the gastric effluent, and the pancreatic enzymes and bile aid in the digestion of starch, proteins, and lipids.

Adaptations

Some characteristics of the vertebrate digestive system are common to all classes of vertebrates and appear to have been conserved in their evolution. Included are most of the enzymes responsible for digestion of carbohydrates, proteins, and lipids, and the mechanisms for absorption of nutrients by the midgut or small intestine. However, the headgut and hindgut show a variety of adaptations to the diet, the environment, or physiological characteristics of the animal.

Many adaptations of the mouthparts and feeding practices to the diet are seen among species in all classes of vertebrates. Prolonged retention of digesta by the hindgut or large intestine for the recovery of urinary and digestive fluids appears to have first developed in the adult amphibians. This resulted in the production of large populations of indigenous bacteria, whose digestive end products aid in the resorption of fluids and can provide additional nutrients.

Further adaptations of the hindgut led to the evolution of herbivores, which have access to a much wider range of food. The major advances in the digestive system of mammals were the development of an extremely efficient masticatory apparatus and, in some species, a more complex and voluminous large intestine or stomach. These advances account for the marked success of the mammalian herbivores and suggest that the success of the herbivorous dinosaurs may have been based on similar developments.

The digestive systems of different species are designed for different diets. The digestive system of most carnivores is designed for the intermittent consumption of high-protein diets, often at intervals of 12–18 h in many mammalian carnivores and longer periods of time for cold-blooded species. The herbivore digestive system is designed for the almost continuous consumption of large volumes of plant material. When left to their normal diet, horses and elephants may spend 18 h a day foraging for food, and ruminants spend 8 h grazing and an additional 8 h remasticating their cud.

A number of the digestive diseases of herbivores can be attributed to the feeding of high-concentrate, low-fiber diets. High levels of rapidly fermentable starch can result in a fulminating production of short-chain fatty acids and the gases CO_2 and methane (CH_4) in the forestomach of cattle and the large intestine of horses. The short-chain fatty acids lower the pH of gut contents, resulting in the growth of lactobacilli and the production of lactic acid. The high levels of short-chain and lactic acids can produce a hypertonic digesta and systemic dehydration, and their absorption can damage the gut epithelium, inhibit gut motility, and result in systemic acidosis. Excessive gas production can distend and displace the gut and inhibit its motility. This syndrome is responsible for a number of diseases involving the forestomach of ruminants and for the high incidence of torsion, impaction, and colic in the large intestine of horses on high-concentrate diets. Although humans are omnivores, the much greater prevalence of cancer and other diseases of the colon on the high-protein, high-carbohydrate, high-lipid, low-fiber diets of affluent Western societies suggests that the human large intestine is subject to similar problems.

The nutritional requirements of vertebrates are provided by an extremely complex system with a variety of different and often exotic adaptations to the diet, habitat, and other physiological characteristics of a species. Studies of these adaptations have provided important contributions to understanding of basic physiological mechanisms. They also provide information needed for the maintenance of domesticated and captive (zoo) animals and survival of endangered species. The extinction of vertebrates at a rate of approximately one species each year could be prevented or at least delayed by a better understanding of these species and their contributions to the beauty and balance of nature. C. Edward Stevens

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Digital computer

A device that processes numerical information; more generally, any device that manipulates symbolic information according to specified computational procedures. The term digital computer—or simply, computer—embraces calculators, computer workstations, control computers (controllers) for applications such as domestic appliances and industrial processes, data-processing systems, microcomputers, microcontrollers, multiprocessors, parallel computers, personal computers, network servers, and supercomputers. *See* CALCULATORS; DIGITAL CONTROL; MICROCOMPUTER; PROGRAMMABLE CONTROLLERS; SUPERCOMPUTER.

A digital computer is an electronic computing machine that uses the binary digits (bits) 0 and 1 to represent all forms of information internally in digital form. Every computer has a set of instructions that define the basic functions it can perform. Sequences of these instructions constitute machine-language programs that can be stored in the computer and used to tailor it to an essentially unlimited number of specialized applications. Calculators are small computers specialized for mathematical computations. General-purpose computers range from pocket-sized personal digital assistants (notepad computers), to medium-sized desktop computers (personal computers and workstations), to large, powerful computers that are shared by many users via a computer network. The vast majority of digital computers now in use are inexpensive, special-purpose microcontrollers that are embedded, often invisibly, in such devices as toys, consumer electronic equipment, and automobiles. *See* BIT; COMPUTER PROGRAMMING; EMBEDDED SYSTEMS.

The main data-processing elements of a computer reside in a small number of electronic integrated circuits (ICs) that form a microprocessor or central processing unit (CPU). Electronic technology allows a basic instruction such as "add two numbers" to be executed many millions of times per second. Other electronic devices are used for program and data storage (memory circuits) and for communication with external devices and human users (input-output circuits). Nonelectronic (magnetic, optical, and mechanical) devices also appear in computers. They are used to construct input-output devices such as keyboards, monitors (video screens), secondary

memories, printers, sensors, and mechanical actuators.

Digital System Basics

The two fundamental ways of representing information are analog and digital. Analog variables take on continuous values that are not separated by gaps, for example, distance measurements on a scale. Digital variables employ a fixed set of noncontinuous or discrete symbols (digits). Analog computers process continuous quantities; a slide rule is an early example. Such devices have been superseded by digital computers, which process information encoded in digital form; the terms computer and digital computer are now largely synonymous. *See* ANALOG COMPUTER; COMPUTER.

Binary codes. The decimal number system preferred by humans employs 10 different digits. The simpler binary or two-valued number system is better suited to computers (see **table**). The bits 0 and 1 have various physical realizations, for example, the position of a switch (off = 0 and on = 1) or an electrical voltage level ($V_{low} = 0$ and $V_{high} = 1$). Modern computers employ both of these concepts by using transistors as high-speed electronic switches to control binary voltage signals. *See* NUMBERING SYSTEMS; TRANSISTOR.

Information is stored and processed by computers in fixed-sized units called words. Common word sizes are 8, 16, 32, and 64 bits. Four-bit words can be used to encode the first 16 integers (see table). By increasing the word size, the number of different items that can be represented and their precision can be made as large as desired. A common word size in personal computers is 32 bits, which allows $2^{32} = 4{,}294{,}967{,}296$ distinct numbers to be represented.

Computer words can represent many different forms of information, not just numbers. For example, 8-bit words called characters or bytes are used to encode text symbols (the 10 decimal digits, the 52 upper- and lowercase letters of the English alphabet, and punctuation marks). A widely used code of this type is ASCII (American Standard Code for Information Interchange). Visual information can be reduced to black and white dots (pixels) corresponding to 0's and 1's. Audio information can be digitized by mapping a small element of sound into a binary word; for example, a compact disk (CD) uses several million 16-bit words to store an audio recording. Logical quantities encountered in reasoning or decision making can be captured by associating 1 with true and 0 with false. Hence, most forms of information are readily reduced to a common, numberlike binary format suitable for processing by computer. *See* COMPACT DISK; IMAGE PROCESSING; LOGIC.

Integers from 0 to 15 in decimal and binary form

Number	Decimal code	Binary code
Zero	00	0000
One	01	0001
Two	02	0010
Three	03	0011
Four	04	0100
Five	05	0101
Six	06	0110
Seven	07	0111
Eight	08	1000
Nine	09	1001
Ten	10	1010
Eleven	11	1011
Twelve	12	1100
Thirteen	13	1101
Fourteen	14	1110
Fifteen	15	1111

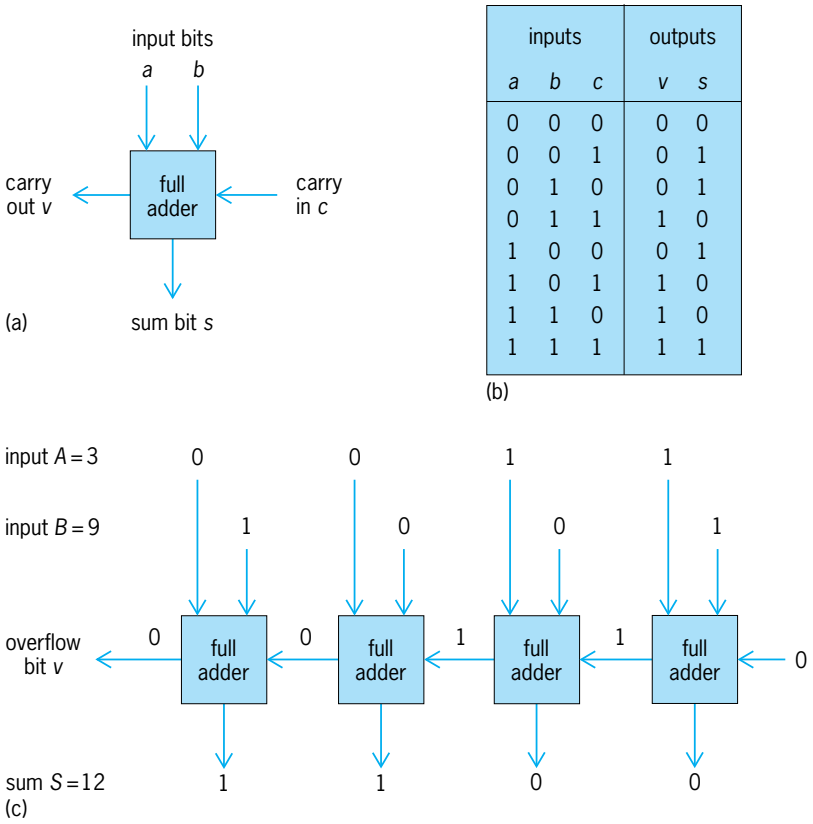

Fig. 1. Logic components. (*a*) Full adder. (*b*) Behavior of a full adder. (*c*) Four-bit adder circuit composed of four full adders.

Logic components. The operation of a digital computer can be viewed at various levels of abstraction, which are characterized by components of different complexity. These levels range from the low, transistor level seen by an electronic circuit designer to the high, system level seen by a computer user. A useful intermediate level is the logic level, where the basic components process individual bits. An example of such a small logic component is the full adder (**Fig. 1***a*), which computes the sum of three bits, *a*, *b*, and *c*. For example, if the input values are $a = b = c = 1$, the output of the full adder is $vs = 11$, denoting three. The behavior of a full adder is specified by a truth table (Fig. 1*b*). By linking four copies of a full adder (Fig. 1*c*), a 4-bit adder is obtained, that is, a circuit that computes the sum of two 4-bit numbers. This adder serves as a component of a larger logic circuit, and is represented by a single box symbol in the diagram of that circuit (**Fig. 2**). The adder's input-output lines, *A*, *B*, and *S*, are grouped into word-transmitting lines called buses. By using

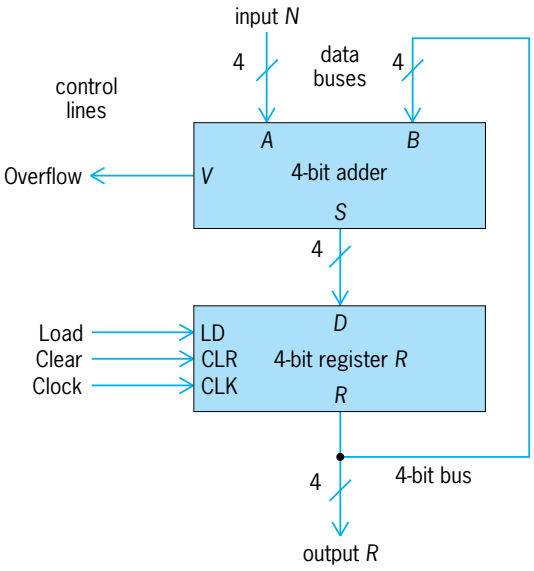

Fig. 2. Four-bit accumulator circuit.

other basic components called gates, logic circuits can be constructed to perform many useful operations. *See* LOGIC CIRCUITS.

A register is a logic component intended to store a word of information. The current contents or state of a register, R, for storing 4-bit words (Fig. 2) appear on its 4-bit output bus, also called R. The register has a 4-bit input bus, D, on which new data are placed for entry into R. It also has several control lines, two of which, Load and Clear, are used to change the state of R. The third control line, Clock, is for timing purposes, and synchronizes the actions of R with those of other registers. Setting the Load control line to 1 causes the current data on D to be entered (loaded) into R and stored there; if Load = 0, the load operation does not occur. Setting the Clear control line to 1 changes the state of R to zero. The precise times at which the register responds to Load and Clear are determined by the signals applied to the Clock line.

An accumulator is a circuit that combines an adder and a register (Fig. 2). It supports a basic type of add instruction found in almost all computers. This instruction is represented symbolically by ADD R, N, which is interpreted as follows: Add the external data, N, to the current contents of the accumulator register, R, and make the resulting sum the new contents of R. Another common instruction, INC R, meaning increment the contents of register R by one, can be implemented by the accumulator by first making $N = 1$ and then executing ADD R, N.This causes the accumulator to behave as a binary counter. *See* DIGITAL COUNTER.

Instructions that make decisions can also be realized by logic circuits. For example, if an answer is needed to the question "Is $R = 0$?" it can be obtained from the 4-bit accumulator by setting $N = 15$, the largest permissible 4-bit number, and then executing the instruction ADD R, N. If $R = 0$, the new value $R + N$ in the register is $0 + 15 = 15$, and the adder's output signal Overflow is 0. If R is greater than 0, then $R + N$ is a 5-bit number between 16 and 30, so the adder overflows, making Overflow = 1. Hence, the Overflow signal, in conjunction with execution of the appropriate form of the ADD instruction, answers the given question.

System organization. An accumulator is a digital system that constitutes a simple processor capable of executing a few instructions. By introducing more data-processing circuits and registers, as well as control circuits for a larger set of instructions, a practical, general-purpose processor can be constructed. Such a processor forms the "brain" of every computer, and is referred to as its central processing unit. A CPU implemented on a single integrated-circuit chip is called a microprocessor. Central processing units have evolved steadily from the simple accumulatorlike processors used in the first generation of electronic computers in the 1940s, to complex microprocessors containing millions of transistors and capable of executing several hundred different instruction types. *See* MICROPROCESSOR.

A typical computer program is too large to store in the CPU, so another component called the main memory is used to store a program's instructions and associated data while they are being executed (**Fig. 3**). Main memory consists of high-speed integrated circuits designed to allow storage and retrieval of information one word at a time. All words in main memory can be accessed with equal ease; hence this is also called a random-access memory (RAM).

A computer program is processed by loading it into main memory and then transferring its instructions and data one word (or a few words) at a time to the CPU for processing. Hence, there is a continual flow of instructions and data words between the CPU

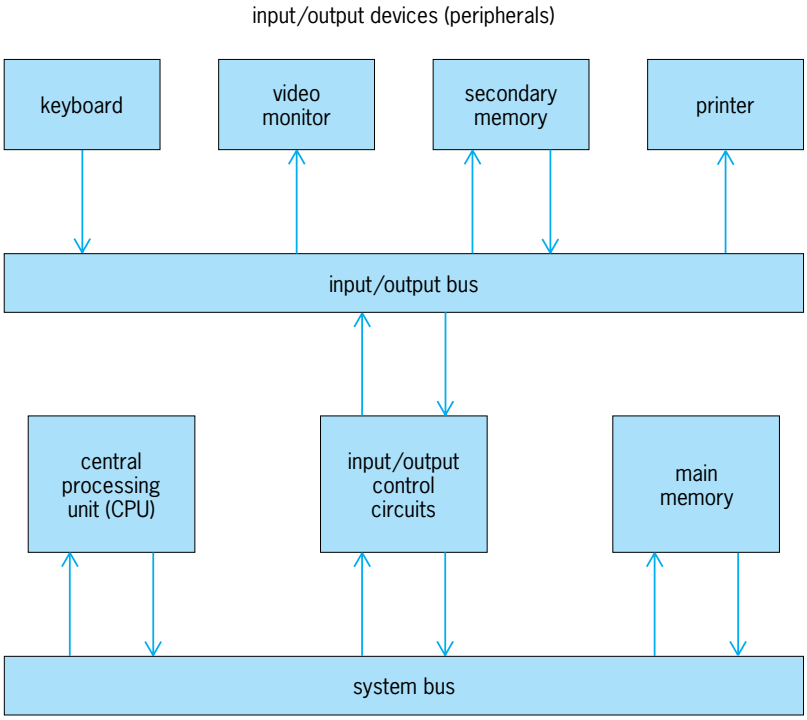

Fig. 3. General organization of a computer.

and its main memory. As millions of words must be transferred per second, a high-speed communication link is needed between the CPU and main memory. The system bus (Fig. 3) fills this role.

A computer has input-output (I/O) control circuits and buses to connect it to external input-output devices (also called peripherals). Typical input-output devices are a keyboard, which is an input device, and a printer, which is an output device. Because most computers need more storage space than main memory can supply, they also employ secondary memory units which form part of the computer's input-output subsystem. Common secondary memory devices are hard disk drives, flexible (floppy) disk drives, and magnetic tape units. Compared to main memory, secondary memories employ storage media (magnetic disks and tapes) that have higher capacity and lower cost. However, secondary memories are also significantly slower than main memory. *See* COMPUTER PERIPHERAL DEVICES; COMPUTER STORAGE TECHNOLOGY.

Central Processing Unit

The CPU contains the circuits needed to fetch instructions from main memory and execute them. Instructions are stored in consecutive memory locations or addresses, in approximately the order in which the instructions must be executed. Hence, all the CPU needs to keep track of instructions is a register that can store instruction addresses and can have its contents incremented. This register is called the program counter. Each time the CPU fetches an instruction word from memory, it increments the program counter to obtain the address of the next instruction it needs (except when branching occurs, as discussed below).

The basic operations that a computer performs are ultimately determined by the machine instruction set that is built into its CPU. All the computer's programs have to be translated into sequences of instructions drawn from this instruction set before the programs can be executed. Most programs are written in high-level programming languages such as C, Cobol, Fortran, and Pascal. The translation of such programs into executable machine-language form is carried out automatically by programs called compilers. *See* PROGRAMMING LANGUAGES.

Every possible computational task can be programmed by using a small number of different types of machine instructions, because a complex operation can be broken down into sequences of simple operations. For example, multiplication can be implemented by a sequence of additions. Thus, while some inexpensive computers have no multiply instructions, that is, they lack the logic circuits to execute multiply in one step, all can be programmed to execute a multistep multiplication algorithm. Multiplication is said to be implemented in software rather than hardware in such computers.

Instruction and data formats. Machine instructions are formatted as binary words with several parts (fields) that define the actions to be taken by the CPU when the instruction is executed. The key field is the opcode (operation code), OP, which specifies the main operation to be performed: for example, "load a 32-bit word from memory to a CPU register," or "add two 64-bit integers." The opcode field is followed by several operand address fields (two are typical), such as OP ADR1,. . . , ADR*k*. (This symbolic, nonbinary format for instructions is known as assembly language, and its format varies from CPU to CPU.) The operand addresses ADR1,. . . ,ADR*k* specify data items stored in various locations: in the instruction itself, in CPU registers, in main memory locations, or in registers associated with input-output control circuits. For example, an instruction of the form SUB R7, #13 specifies that the constant 13 is to be subtracted from the contents of the register named R7. Register R7 serves as both a data source and data destination in this two-address instruction format. The one-address instruction LOAD 102678 transfers the contents of the memory location with the indicated address 102678 to an implicit register in the CPU intended for such operations. Programs can be written in assembly languages of the above type; but as noted above, most programming is now done with higher-level languages that are more expressive and easier to use.

Numerical operands have two main formats: fixed-point to represent integers and floating-point to represent real numbers. Fixed-point numbers take the form $b_{n-1}b_{n-2}\cdots b_2b_1b_0$, where b_i is 0 or 1 and has the numerical weight 2^i (see table). The leftmost position of a number word can be reserved to represent the number's sign, with 0 denoting plus and 1 denoting minus. A floating-point number consists of a pair of fixed-point numbers M and E, which typically denote the number $M \times 2^E$. Nonnumerical data usually take the form of variable-length character strings encoded in a standard code such as ASCII.

Instruction types. The choice of the instructions to include in an instruction set is governed by several factors: the usefulness of the instructions, the cost of the hardware to implement them, and their compatibility with the instructions of other computers. Because their data are physically closer to the CPU's data-processing circuits, instructions that act on CPU registers execute faster than instructions of the same type whose operands are in memory.

The operations performed by the most common instruction types are summarized below:

1. *Load.* Copy an operand from a main memory location to a CPU register.
2. *Store.* Copy an operand to a main memory location from a CPU register.
3. *Move register.* Copy an operand from one CPU register to another.
4. *Add.* Compute the sum of two operands stored in CPU registers.
5. *Subtract.* Compute the difference of two operands stored in CPU registers.
6. *Not.* Compute the logical complement of an operand in a CPU register by changing its 1's to 0's and its 0's to 1's.
7. *Clear.* Change every bit of a CPU register to 0.
8. *Branch.* Replace the contents of the program

counter by the instruction's address field. (This is an unconditional branch instruction.)

9. *Branch on zero.* If the contents of a specified CPU register are zero, replace the contents of the program counter by the instruction's address field. (This is a conditional branch instruction.)

No explicit instructions are needed for input-output operations if input-output devices share with main memory the available memory addresses. This is known as memory-mapped input-output, and allows load and store instructions to be used to transfer data between the CPU and input-output devices.

In general, a computer's instruction set should include a selection of instructions of the following three types: (1) Data-transfer instructions that move data unchanged between the CPU, main memory, and input-output devices. (2) Data-processing instructions that perform numerical operations such as add, subtract, multiply, and divide, as well as nonnumerical (logical) operations, such as not, and, exclusive-or, and shift. (3) Program-control instructions that can change the order in which instructions are executed, for example branch, branch-on-zero, call procedure, and return from procedure.

Internal organization. The instruction unit (I unit) of a CPU (**Fig. 4**), also called the program control unit, is responsible for fetching instructions from main memory, using the program counter as the instruction address register. The opcode of a newly fetched instruction *I* is placed in the instruction register. The opcode is then decoded to determine the sequence of actions required to execute *I*. These may include the loading or storing of data assigned to main memory, in which case the I unit computes all needed addresses and issues all needed control signals to the CPU and the system bus. Data are processed in the CPU's execution unit (E unit), also called the datapath, which contains a set of registers used for temporary storage of data operands, and an arithmetic logic unit (ALU), which contains the main data-processing circuits.

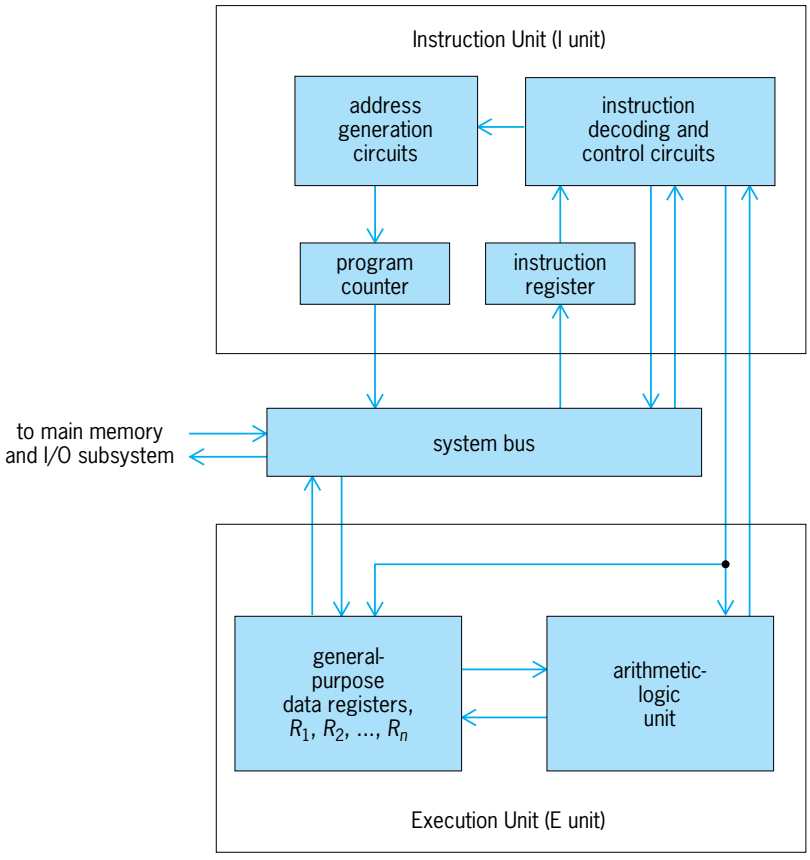

Fig. 4. Internal organization of a CPU.

Computer Spectrum

A major theme in the evolution of digital computers has been to find ways to increase the computing speed or performance of a computer at low cost. Computers have undergone a long period of development, with major advances stimulated by new technologies for manufacturing processors and memories.

Development. The modern digital computer has its origins in the work of C. Babbage, who attempted unsuccessfully to build a general-purpose, programmable computer in the mid-1800s. Around 1940, such computers were successfully built in the United States and Europe by using electronic vacuum tubes as the basic switching devices. These led to the first generation of computers. The invention of the transistor in 1947, which replaced the vacuum tube, gave rise to a second generation of smaller, faster, and more reliable computers. Most early computers were expensive, room-sized mainframe machines. *See* VACUUM TUBE.

In the mid-1960s, the first series of mainframe computers to distinguish a computer's architecture from its implementation was introduced. The architecture of a computer series is defined as a specification covering the instruction set, and the high-level structure and behavior of all members of the series. Different models in the series differ in their low-level (implementation) details. Another innovation of the 1960s was the minicomputer, a small (desk-sized) and relatively inexpensive computer with limited hardware and software facilities. Integrated transistor circuits also began to displace discrete transistor circuits, further reducing computer size and cost. A milestone occurred in 1971 with the introduction of the microprocessor. The density and speed of integrated-circuit technology grew rapidly, eventually making it possible to incorporate into a single chip most of the features of mainframe computers. As the manufacturing cost per integrated circuit is fairly constant, these developments have meant a steady improvement in the performance/cost ratio of computers. *See* COMPUTER SYSTEMS ARCHITECTURE; INTEGRATED CIRCUITS.

By 1980, single-chip microprocessors and microcomputers were being mass-produced in several varieties. The simplest microcomputers were being used as microcontrollers for dedicated applications. The more powerful microprocessors began to appear in inexpensive general-purpose computers, a development that produced the personal computer. By 1990, all computers could be placed in a few families

defined by certain microprocessor and operating-system standards. *See* OPERATING SYSTEM.

Performance measures. A simple indicator of a CPU's performance is the frequency f of its central timing signal (clock), measured in millions of clock signals issued per second or megahertz (MHz). The clock frequency depends on the integrated-circuit technology used; frequencies of several hundred megahertz are achievable with current technology. Each clock signal triggers execution of a basic instruction such as a fixed-point addition; hence, the time required to execute such an instruction (the clock cycle time) is $1/f$ microseconds. Complex instructions like multiplication or operations on floating-point numbers require several clock cycles to complete their execution. Another measure of CPU performance is the (average) instruction execution rate, measured in millions of instructions per second (MIPS).

Instruction execution time is strongly affected by the time to move instructions or data between the CPU and main memory. The time required by the CPU to access a word in main memory is typically about five times longer than the CPU's clock cycle time. This disparity in speed has existed since the earliest computers despite efforts to develop memory circuits that would be fast enough to keep up with the fastest CPUs. Maximum performance requires the CPU to be supplied with a steady flow of instructions that need to be executed. This flow is disrupted by branch instructions, which account for 20% or more of the instructions in a typical program.

To deal with the foregoing issues, various performance-enhancing features have been incorporated into the design of computers. The communication bottleneck between the CPU and main memory is reduced by means of a cache, which is a special memory unit inserted between the two units. The cache is smaller than main memory but can be accessed more rapidly, and is often placed on the same integrated-circuit chip as the CPU. Its effect is to reduce the average time required by the CPU to send information to or receive information from the memory subsystem. Special logic circuits support the complex flow of information among main memory, the cache, and the registers of the CPU. However, the cache is largely invisible to the programs being executed.

The instruction execution rate can be increased by executing several instructions concurrently. One approach is to employ several E units that are tailored to different instruction types. Examples are an integer unit designed to execute fixed-point instructions and a floating-point unit designed for floating-point instructions. The CPU can then execute a fixed-point instruction and a floating-point instruction at the same time. Processors that execute several instructions in parallel in this way are called superscalar. *See* CONCURRENT PROCESSING.

Another speedup technique called pipelining allows several instructions to be processed simultaneously in special circuits called pipelines. Execution of an instruction is broken into several consecutive steps, each of which can be assigned to a separate stage of the pipeline. This makes it possible for an n-stage E unit to overlap the execution of up to n different instructions. A pipeline processing circuit resembles an assembly line on which many products are in various stages of manufacture at the same time. The ability of a CPU to execute several instructions at the same time by using multiple or pipelined E units is highly dependent on the availability of instructions of the right type at the right time in the program being executed. A useful measure of the performance of a CPU that employs internal parallelism is the average number of clock cycles per instruction (CPI) needed to execute a representative set of programs.

CISCs and RISCs. A software implementation of a complex operation like multiply is slower than the corresponding hardware implementation. Consequently, as advances in IC technology lowered the cost of hardware circuits, instruction sets tended to increase in size and complexity. By the mid-1980s, many microprocessors had instructions of several hundred different types, characterized by diverse formats, memory addressing modes, and execution times. The heterogeneous instruction sets of these complex instruction set computers (CISCs) have some disadvantages. Complex instructions require more processing circuits, which tend to make CISCs large and expensive. Moreover, the decoding and execution of complex instruction can slow down the processing of simple instructions.

To address the defects of CISCs, a new class of fast computers referred to as reduced instruction set computers (RISCs) was introduced. RISCs are characterized by fast, efficient—but not necessarily small—instruction sets. The following features are common to most RISCs: (1) All instructions are of fixed length and have just a few opcode formats and addressing modes. (2) The only instructions that address memory are load and store instructions; all other instructions require their operands to be placed in CPU registers. (3) The fetching and processing of most instructions is overlapped in pipelined fashion.

An example of a RISC is the PowerPC microprocessor (**Fig. 5**), developed in the early 1990s. It is a family of single-chip, 32-bit microprocessors which share the same RISC architecture. The PowerPC has a set of 32 general-purpose registers for storing integer operands, and 32 more registers for floating-point operands. There are more than 200 distinct instruction types, all one word in length. The only instructions that address memory are load and store. Extensive use is made of pipelines to ensure that most instructions can be executed in one clock cycle.

The PowerPC incorporates several speedup features which are not unique to RISCs. It contains a cache whose size and organization varies with the model; the cache of the model 601 (Fig. 5) has a capacity of 32 kilobytes (2^{15} bytes) and stores both data and instructions. The PowerPC has several E units and so falls into the superscalar category. The model 601 has three E units: an integer execution unit, a floating-point unit, and a branch processing

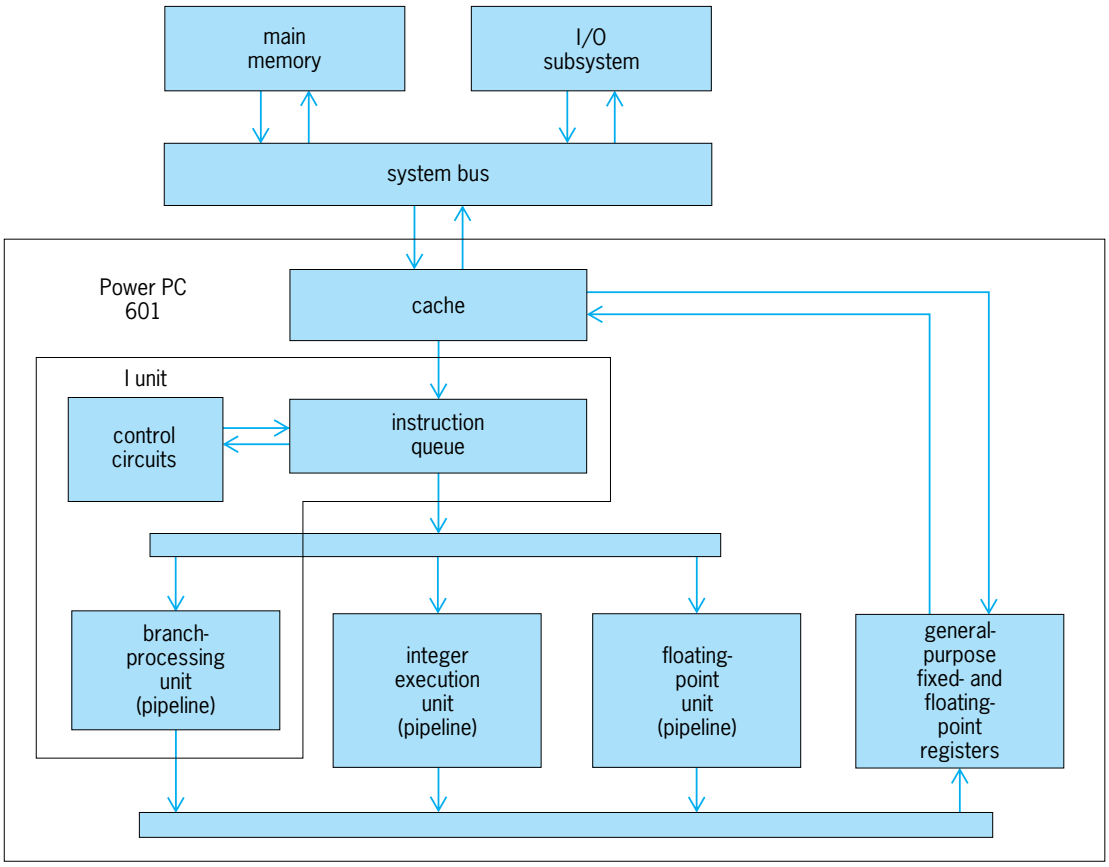

Fig. 5. Organization of the PowerPC computer, model 601.

unit, which allow up to three instructions to complete execution in one clock cycle. The integer unit executes all fixed-point numerical and logical operations, including those associated with load and store instructions. Although part of the CPU's I unit, the branch processing unit is considered an E unit for branch instructions. To keep the three E units supplied with instructions, the I unit can fetch up to eight instructions simultaneously from the cache.

Parallel processing. The degree of parallelism possible with a pipeline is small, less than 10. A parallel processing technique with the potential of achieving very great parallelism is the use of multiple independent computers or CPUs operating in unison. For example, a set of computers in a computer network can be programmed to work concurrently on different parts of the same problem. Such a loosely coupled computer system is useful for computing tasks that can easily be partitioned into independent subtasks, with infrequent communication of results among the subtasks.

Some large-scale computations can be divided into subtasks that require more frequent and rapid exchange of results among the subtasks than is possible with a computer network. To address this interprocessor communication problem, computers have been built whose CPUs are tightly coupled so they can access one another's data rapidly; these are called multiprocessors. The task of writing programs for multiprocessors is not as well understood as the corresponding problem for a single-processor computer. Nevertheless, powerful multiprocessors employing many inexpensive microprocessors as their CPUs are available commercially. They are increasingly used in applications that have extremely high performance requirements. *See* MULTIPROCESSING.

John P. Hayes

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Digital control

The use of digital or discrete technology to maintain conditions in operating systems as close as possible to desired values despite changes in the operating environment. Traditionally, control systems have utilized analog components, that is, controllers which generate time-continuous outputs (volts, pressure, and so forth) to manipulate process inputs and which operate on continuous signals from instrumentation measuring process variables (position, temperature, and so forth). In the 1970s, the use of discrete or logical control elements, such as fluidic components, and the use of programmable

logic controllers to automate machining, manufacturing, and production facilities became widespread. In parallel with these developments was the accelerating use of digital computers in industrial and commercial applications areas, both for logic-level control and for replacing analog control systems. The development of inexpensive mini- and microcomputers with arithmetic and logical capability orders of magnitude beyond that obtainable with analog and discrete digital control elements resulted in the rapid substitution of conventional control systems by digital computer–based ones. With the introduction of microcomputer-based control systems into major consumer products areas (such as automobiles and video and audio electronics), it became clear that the digital computer would be widely used to control objects ranging from small, personal appliances and games up to large, commercial manufacturing and production facilities. Hence the focus of this discussion will be on computer-based control systems. *See* MICROCOMPUTER; MICROPROCESSOR; PROGRAMMABLE CONTROLLERS.

Computer/process interface. The object that is controlled is usually called a device or, more inclusively, a process. A characteristic of any digital control system is the need for a process interface to mate the digital computer and process, to permit them to pass information back and forth (**Fig. 1**).

Digital control information. Measurements of the state of the process often are obtained naturally as one of two switch states; for example, a part to be machined is in position (or not), or a temperature is above (or below) the desired temperature. Control signals sent to the process often are expressed as one of two states as well; for example, a motor is turned on (or off), or a valve is opened (or closed). Such binary information can be communicated naturally to and from the computer, where it is manipulated in binary form. For this reason the binary or digital computer–process interface usually is quite simple: a set of signal-conditioning circuits for each measured or controlled signal and a set of registers to transfer the bits of digital information in each direction. Each register usually contains the same number of bits as would be manipulated and stored within the digital computer.

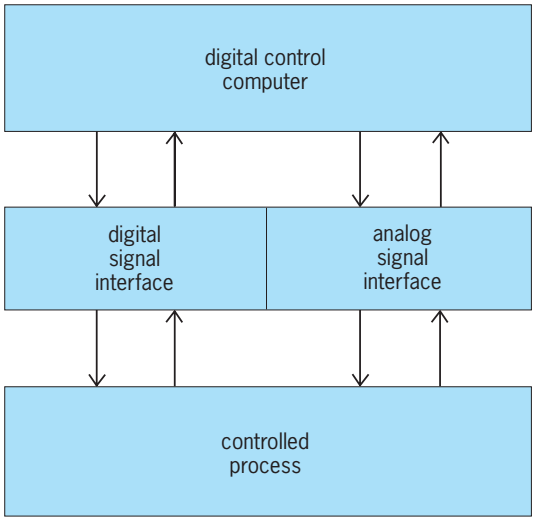

Fig. 1. Generalized digital computer control system.

Analog control information. Process information also must be dealt with in analog form; for example, a variable such as temperature can take on any value within its measured range or, looked at conceptually, it can be measured to any number of significant figures by a suitable instrument. Furthermore, analog variables generally change continuously in time. Digital computers are not suited to handle arbitrarily precise or continuously changing information; hence, analog process signals must be reduced to a digital representation (discretized), in terms of both magnitude and time, to put them into a useful digital form.

The magnitude discretization problem most often is handled by transducing and scaling each measured variable to a common range, then using a single conversion device—the analog-to-digital converter—to put the measured value into digital form. An analog-to-digital converter suitable for measurement and control purposes typically will convert signals in the range −10 to +10 V direct current, yielding an output with 12 to 16 bits of precision in 10 to 50 microseconds. A multiplexer often is used to allow a number of analog inputs to be switched into a single analog-to-digital converter. High-level signals (on the order of volts) can be switched by a solid-state multiplexer; low-level signals (on the order of millivolts, from strain gages or thermocouples) require mechanical relays followed by an amplifier to boost the signal to an acceptable input level for the analog-to-digital converter. Microcomputers are now manufactured that contain integral analog conversion circuitry for several channels on the processor chip. *See* ANALOG-TO-DIGITAL CONVERTER.

Discretization in time requires the computer to sample the signal periodically, storing the results in memory. This sequence of discrete values yields a "staircase" approximation to the original signal (**Fig. 2**), on which control of the process must be based. Obviously, the accuracy of the representation can be improved by sampling more often, and many digital systems simply have incorporated traditional analog control algorithms along with rapid sampling. However, newer control techniques make fundamental use of the discrete nature of computer input and output signals. Analog outputs from a computer most often are obtained from a digital-to-analog converter, a device which accepts a digital output from the computer, converts it to a voltage in several microseconds, and latches (holds) the value until the next output is converted. Usually a single digital-to-analog converter is used for each output signal. *See* DIGITAL-TO-ANALOG CONVERTER.

Real-time computing. In order to be used as the heart of a control system, a digital computer must be capable of operating in real time. Except for very simple microcomputer applications, this feature implies that the machine must be capable of handling interrupts, that is, inputs to the computer's internal

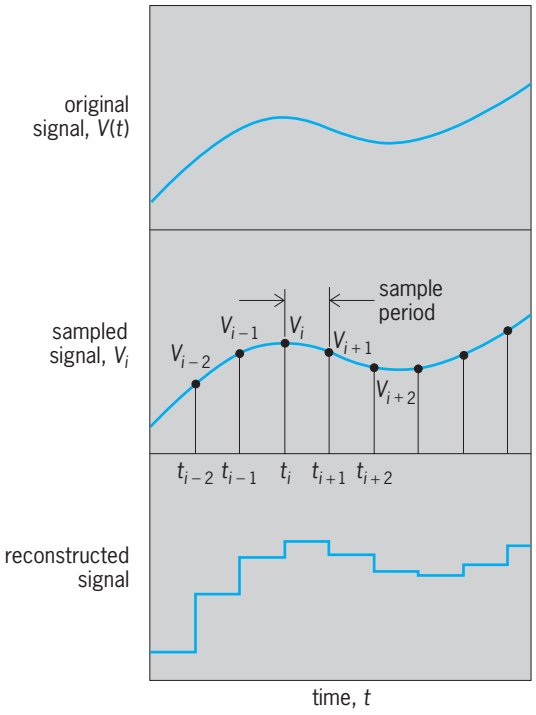

Fig. 2. Discretization in time of an analog signal.

control unit which, on change of state, cause the computer to stop executing some section of program code and begin executing some other section. Using its extraordinary computational abilities (on the order of a million instructions per second), the computer, which basically is a sequential or serial processor, can be made to appear to perform operations in parallel by proper design of its hardware and executive software. By attaching a very accurate oscillator (a so-called real-time clock) to the interrupt line, the computer can be programmed to keep track of the passage of real time and, consequently, to schedule process sampling and control calculations on a periodic basis.

The requirements of real-time computing also imply that the computer must respond to interrupts from the process. Thus a key process variable may be used to trigger an interrupt when it exceeds preset limits. The computer would be programmed to service such an interrupt immediately, taking whatever control action would be necessary to bring the variable back within limits. The ability to initiate operations on schedule and to respond to process interrupts in a timely fashion is the very basis of real-time computing; this feature must be available in any digital control system.

Programming considerations. Much of the programming of computer control systems is done in a high-level language such as BASIC, FORTRAN, C, or Ada; however, many microcomputer applications are carried out in machine (assembly) language. In either case the programmer must utilize program routines which access the devices in the process interface, for example, fetch the contents of a particular digital input register so as to check the status of some process digital element, or write out a digital result to a particular digital-to-analog converter channel. Additionally, the programmer will have the capability to schedule operations periodically or at particular times. *See* PROGRAMMING LANGUAGES.

Computer control systems for large or complex processes may involve complicated programs with many thousands of computer instructions. Several routes have been taken to mitigate the difficulty of programming control computers. One approach is to develop a single program which utilizes data supplied by the user to specify both the actions to be performed on the individual process elements and the schedule to be followed. Such an executive program, supplied by a control system vendor, might be utilized by a variety of users; and this approach often is taken for relatively standardized operations such as machining and sequential processing (manufacturing).

Another approach is to develop a rather sophisticated operating system to supervise the execution of user programs, scheduling individual program elements for execution as specified by the user or needed by the process. The multiple program elements, called tasks when used with a multitasking operating system, or called programs with a multiprogramming system, will be scheduled individually for execution by the operating system as specified by the user (programmer) or needed by the process. A simple example is given in **Fig. 3** for three tasks—a data logging task which might be scheduled every 30 s, a process alarm task which would be executed whenever a process variable exceeds limits and causes an interrupt, and an operator communication task which would be executed whenever the operator strikes a key on the console. The operating system scheduler would resolve conflicts, caused by two or three tasks needing to execute simultaneously, on the basis of user-supplied priorities; presumably the alarm task would have the highest priority here. *See* COMPUTER PROGRAMMING; SOFTWARE ENGINEERING.

Control algorithms. Many applications, particularly machining, manufacturing, and batch processing, involve large or complex operating schedules. Invariably, these can be broken down into simple logical sequences, for example, in a high-speed bottling operation the bottle must be in position before the filling line is opened. Hence the control program reduces to a set of interlocked sequential operations. Some applications—in the chemical process industries, in power generation, and in aerospace areas—require the use of traditional automatic control algorithms.

Automatic control algorithms fall into two major categories: feedforward techniques, where process disturbances which would affect the controlled variable are measured and their effect canceled out or compensated by appropriate manipulation of a process input variable; and feedback techniques. In feedback control, which makes up the vast majority of applications, the controlled variable itself is measured and subtracted from a reference (set-point) variable equal to its desired value. The resulting deviation is

the input to the controller, which then manipulates a process input variable through the final control element. If the controller is designed correctly, it should maintain the deviation quite small despite operator changes to the set point or despite environmental changes which cause disturbances to enter the control loop (**Fig. 4*a***).

Digital implementation of traditional analog algorithms. Traditional algorithms for analog control of processes have been developed over many years. One very important example is the three-mode controller, so called because the algorithm for the controller output U operates proportionally on the deviation ϵ, the time integral of ϵ, and the time derivative of ϵ as indicated in Eq. (1), where U_0 is the constant value that the

$$U = U_0 + K_c\left(\epsilon + \frac{1}{\tau_I}\int_0^t \epsilon dt + \tau_D\frac{d\epsilon}{dt}\right) \quad (1)$$

controller output assumes when the controller first is turned on, and K_c, τ_I, and τ_D are the controller proportional gain, and integral and derivative time parameters, respectively. These controller parameters are chosen by the system designer appropriately for each control application.

In the digital implementation of a single-loop feedback control system (Fig. 4*b*), the digital computer implements the controller algorithm by using process information sampled through the analog-to-digital converter and a set point supplied by the operator. The symbolic switches and the subscript on the controller input ϵ_k (obtained from R_k and M_k) in Fig. 4*b* are intended to indicate that the information is obtained by the computer only at discrete sampling times, T time units apart. Similarly, the controller generates an output U_k at each sampling time which it transmits through the digital-to-analog converter to the process's final control element. Implementation of the three-mode algorithm in the digital (discrete) environment is straightforward. In the "velocity form" of the proportional-integral-derivative algorithm, so called because the change in U from its previous value is calculated directly [Eq. (2)], the

$$U_k = U_{k-1} + K_c\left[(\epsilon_k - \epsilon_{k-1}) + \frac{T}{\tau_I}\epsilon_k + \frac{\tau_D}{T}(\epsilon_k - 2\epsilon_{k-1} + \epsilon_{k-2})\right] \quad (2)$$

computer need only store the controller output and two previous values of ϵ in order to calculate U at the next sampling time. In actual application, digital versions of continuous control algorithms often utilize rapid sampling (T is small) so that they respond in much the same way as analog systems do; hence the operational advantages and disadvantages of the two methods are equivalent.

Purely digital algorithms. One disadvantage of the classical feedback algorithm of Eq. (1) or (2) is that it does not function well when the dynamics of the process contain significant time delay, that is, when a change in the process input has no effect on the controlled variable for a period of time known as the dead time. So-called time-delay compensation algorithms have been available in analog form; however, only with the digital computer as controller has the implementation become so inexpensive that such applications can be widely justified. Hence the use of the three-mode algorithm in conjunction with a time-delay compensator is a common digital control approach, particularly for control of

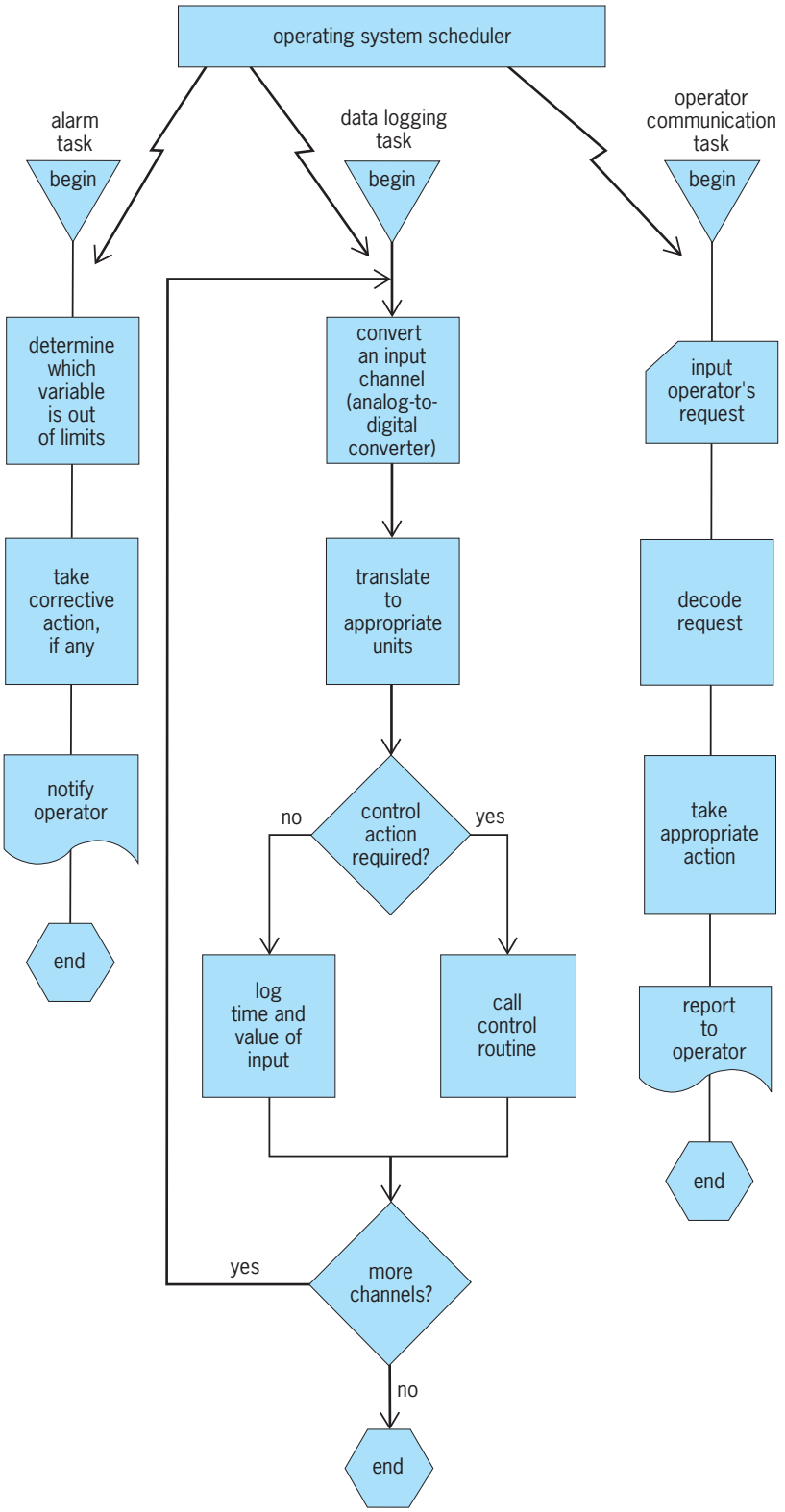

Fig. 3. Example of a multitasked control program.

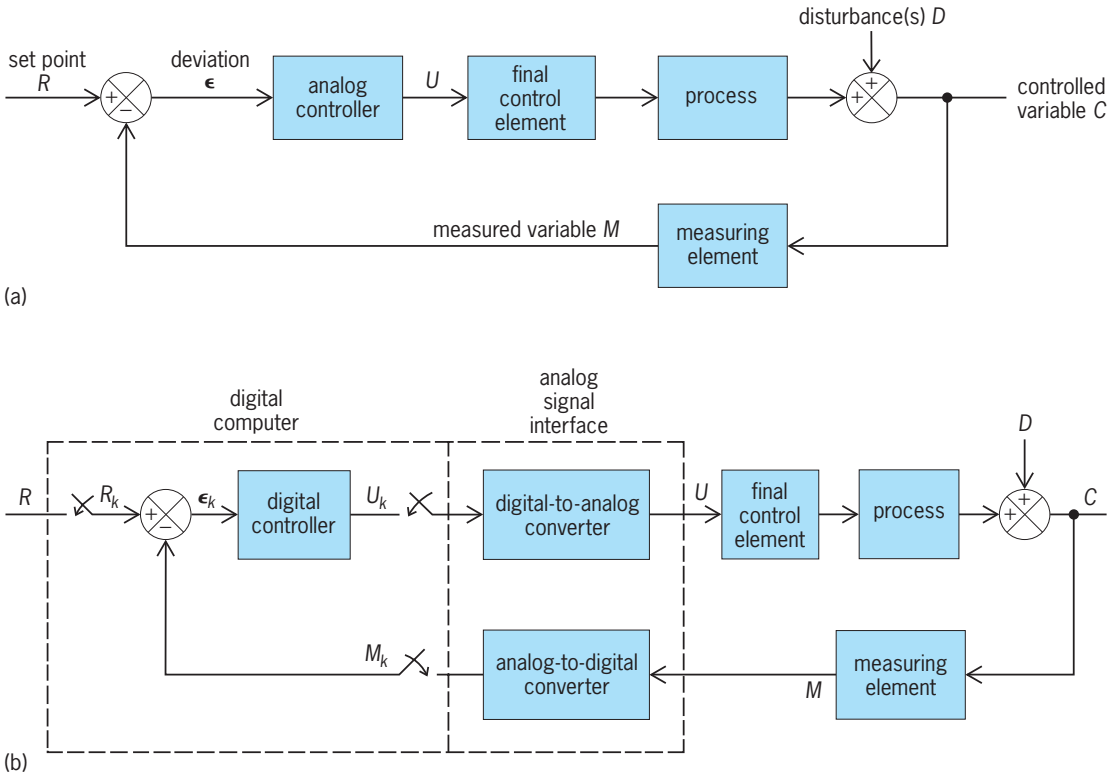

Fig. 4. Single-loop feedback control system. (*a*) Traditional analog implementation. (*b*) Digital implementation.

processes (where time delay is an important component of many, if not most, systems).

Attempts to expand the digital control medium through development of strictly digital control algorithms is an important and continuing trend. Such algorithms typically attempt to exploit the sampled nature of process inputs and outputs, significantly decreasing the sampling requirements of the algorithm (that is, making T as large as possible). For example, in the "deadbeat" algorithm, Eq. (3), the con-

$$U_k = U_{k-1} + \frac{1}{K(N+1)(1-e^{-T/\tau})}(\epsilon_k - e^{-T/\tau}\,\epsilon_{k-1}) \tag{3}$$

troller attempts to follow set-point changes in minimum time. The use of this algorithm presupposes that the process can be modeled approximately as a first-order plus time-delay system, where K and τ are the process gain and time constant and NT is the time delay. Hence a stepwise change in set point will cause the controlled variable to obtain its new, desired value in $N + 1$ sampling instants, the absolute minimum.

The deadbeat algorithm has a number of disadvantages, including its sensitivity to errors in the assumed process model. These often result in undesirable oscillations (ringing). Dahlin's algorithm, Eq. (4), is one of a number of attempts to eliminate

$$U_k = U_{k-1} + \frac{1}{\left\{\begin{array}{c} K[N(1-e^{\lambda T})+1] \\ \times\,(1-e^{-T/\tau})(\epsilon_k - e^{-T/\tau}\epsilon_{k-1}) \end{array}\right\}} \tag{4}$$

the undesirable features of optimal, deadbeat control. In this instance, the use of an empirical tuning parameter, λ, relaxes the requirement of minimal response time, resulting in less ringing in the system response, particularly when the process dynamics change slightly from those assumed in the first-order model.

Dahlin's and the deadbeat algorithms, and other similar algorithms which are designed on the basis of set-point response, often respond more poorly to disturbances than the three-mode algorithm. Hence there is a continuing search for the "best" algorithm—one which is simple, responds well to both set-point changes and disturbances, requires little knowledge of process dynamics and is insensitive to changes in process dynamics, tolerates process constraints and nonlinearities, can be extended to multi-input–multioutput situations, is easy to implement and program, and so forth. Such a universal digital control algorithm likely will not be found, but the search has led in a number of interesting directions, including the development of self-tuning algorithms and the application of time convolution techniques for predictive control strategies. These approaches would not be possible without the digital computer and the ability to design and implement computer control systems. Additional demands are made as well in the analysis of digital control systems: discrete algebraic (difference) equations substitute for differential equations used traditionally to describe processes and analog controllers; and discrete transform techniques (the Z transform) often replace the Laplace transform approach used in

analyzing continuous systems. *See* ALGORITHM; CONTROL SYSTEMS; DIGITAL COMPUTER; INTEGRAL TRANSFORM.

Duncan A. Mellichamp

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Digital counter

A device which, in its simplest form, provides an output that corresponds to the number of pulses applied to its input.

Types. Counters may take many forms: geared mechanisms (water meters, odometers), relays (early telephone switch gear), vacuum tubes (early test instruments), and solid-state semiconductor circuits. This article will stress simple solid-state electronic counter circuits.

Most digital counters operate in the binary number system, since such a system is easily implemented with electronic circuitry. The number of input pulses is represented as a series of binary digits, or bits, where each bit is either a 0, or a 1 (a voltage of 0 or 5 V, usually). *See* NUMBERING SYSTEMS.

Figure 1 shows a four-bit binary counter that will count from 0 to 15; the sixteenth count input causes the counter to return to the 0 state and generate a carry pulse. This action to return to the 0 count with a carry output every sixteenth pulse makes the counter a modulo 16 counter. The four binary outputs Q_D, Q_C, Q_B, and Q_A represent the states of four flip-flops irrespective of whether the count input is 0 or 1. The counter is, therefore, a four-state-variable Moore state machine. Outputs Q_D, Q_C, Q_B, and Q_A are said to to have an 8-4-2-1 "weighting," because if Q_D through Q_A are all 1's, the binary counter output is $1111_2 = 1 \times 2^3 + 1 \times 2^2 + 1 \times 2^1 + 1 \times 2^0 = 8 + 4 + 2 + 1 = 15_{10}$, where the subscripts indicate the base of the number system. In Fig. 1*a*, the counter state-flow diagram is shown. The word "state" refers to the number, the datum, in the counter; this datum is stored in four flip-flop devices, each of which is a memory bit that stores the current value of one binary digit. Each possible state is represented by the numerical output of that state inside a circle. Upon receiving a count pulse, the counter must change state by following an arrow from the present state to the next state. In Fig. 1*b*, a table shows the counter output after a given number of input pulses, assuming that the counter always starts from the 0 state. The counter output is listed in binary, octal, decimal, and hexadecimal. Figure 1*c* shows a block diagram of the counter built with *T* flip-flops, and Fig. 1*d* shows the counter waveforms through time, with a periodic count input. The *T* flip-flop is a device that has either a 0 or a 1 on its *Q* output at all times. When the count input *T* moves from the 1 state to the 0 state, the flip-flop output must change state, from a 0 to a 1 or from a 1 to a 0. The carry output produces a 1-to-0 transition on every sixteenth count input, and therefore a divide-by-16 function.

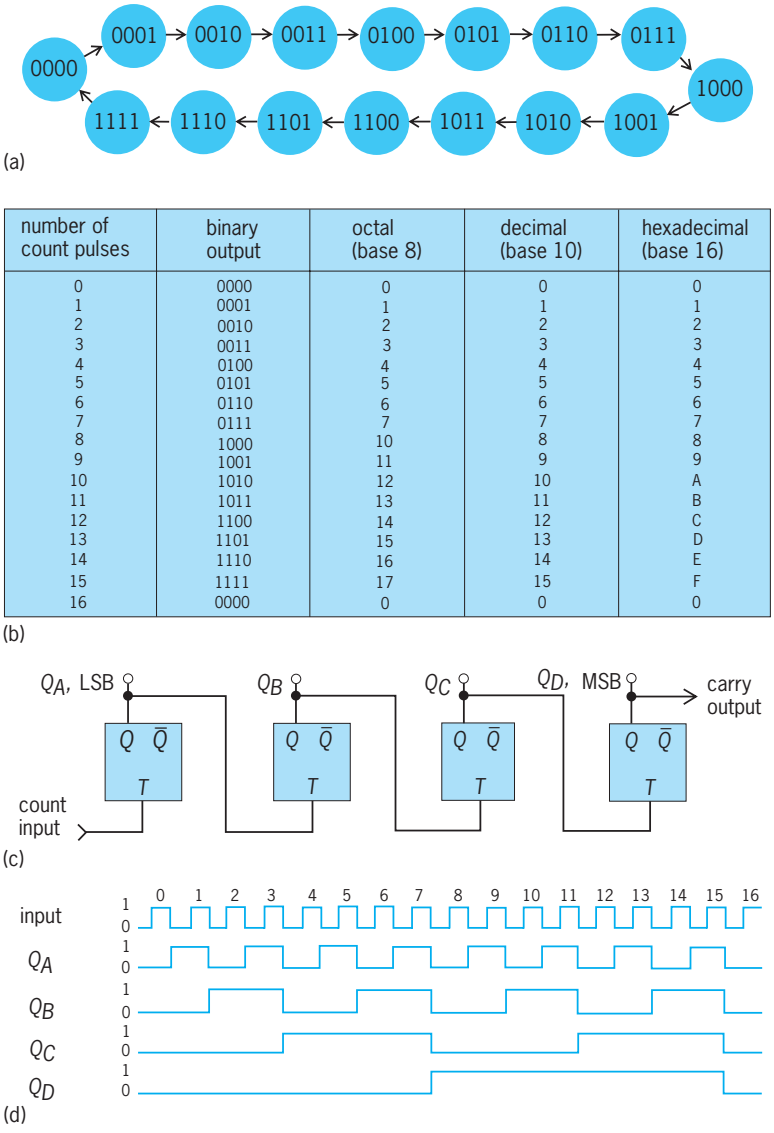

Fig. 1. Four-bit binary counter using trigger flip-flops; this is a Moore machine. (*a*) State-flow diagram. (*b*) Table of the counter output in various number systems. (*c*) Circuit block diagram. LSB = least significant bit; MSB = most significant bit. (*d*) Output waveforms.

The four bits of the counter of Fig. 1 can be grouped together and used to represent a single hexadecimal digit; in Fig. 1*b*, each counter output state represents one hexadecimal digit. A two-digit hexadecimal counter requires two sets of four-bit binary counters, the carry output from the first set of counters driving the count input of the second set of counters.

A decimal counter built from four binary counters is shown in **Fig. 2**. Let four bits of data from the binary counter represent one decimal digit. The counter will work in the same way as the counter of Fig. 1, except that all the flip-flops are reset to the 0 state when the counter moves from the $1001_2 = 9$

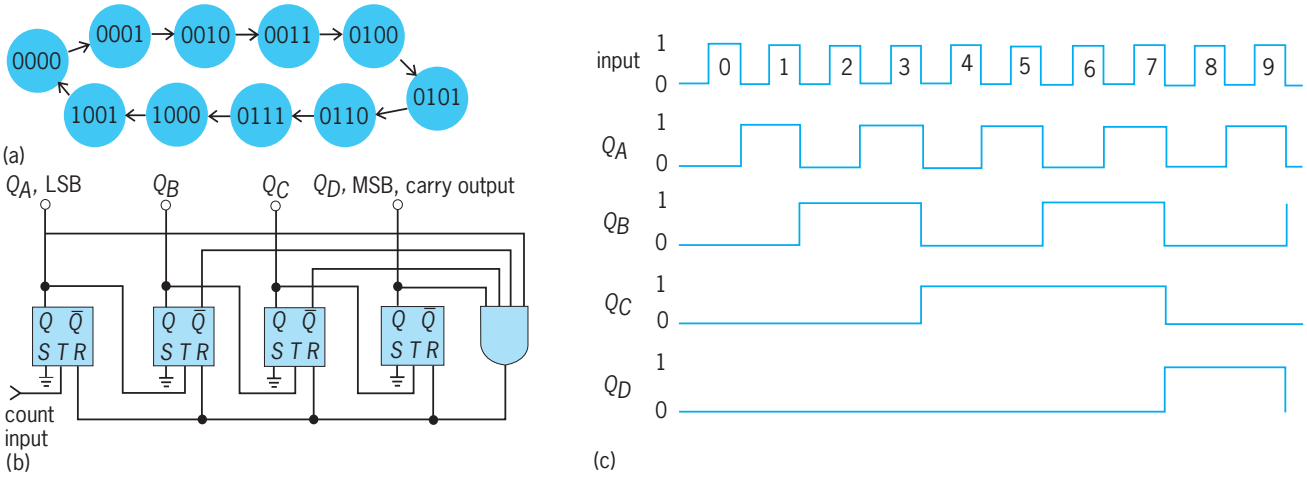

Fig. 2. Four-bit binary counter modified to be a decimal counter. (*a*) State-flow diagram. (*b*) Circuit block diagram. LSB = least significant bit; MSB = most significant bit. (*c*) Output waveforms.

state, instead of advancing to the $1010_2 = 10$ state. Besides the AND gate that is now used to detect the 1001 state of the counter and enable the resets, the circuit block diagram shows a new type of flip-flop. The *"SR"* flip-flop acts like a *T* flip-flop with an additional input that forces the *Q* output to a 1 state when the *S* (set) input is high and the *T* input has had a 1-to-0 transition applied. An *R* (reset) input acts as the *S* input does, except that the *Q* output goes to 0. This example decimal counter has an 8-4-2-1 weighted output that is known as binary coded decimal (B.C.D.). A seven-segment display is easily interfaced to the binary-coded-decimal counter using binary-coded-decimal-to-seven-segment decoder/driver circuits that are widely available.

Applications. Digital counters are found in much modern electronic equipment, especially equipment that is digitally controlled or has digital numeric displays. A frequency counter, as a test instrument or a channel frequency display on a radio tuner, consists simply of a string of decade counters that count the pulses of an input signal for a known period of time, and display that count on a seven-segment display. A digital voltmeter operates by using nearly the same idea, except that the counter counts a known frequency for a period of time proportional to the input voltage. *See* ANALOG-TO-DIGITAL CONVERTER.

A digital watch contains numerous counter/dividers in its large-scale integration (LSI) chip, usually implemented with complementary metal oxide semiconductor (CMOS) technology. A typical watch generates a 32,768-Hz crystal oscillator signal that is divided by 2^{15} (15 *T* flip-flop binary counters) to produce a 1-Hz signal that is counted as seconds by using two decimal counters that reset and produce a carry at every sixtieth count. That carry output is counted as a minutes display, generating another modulo 60 carry pulse for the hours display. The hours counters produce a carry at every twenty-fourth count to run the days counter and display; that counter generates a carry pulse every 28 to 31 days, depending on the month. *See* INTEGRATED CIRCUITS; WATCH.

Digital computers may contain counters in the form of programmable interval timers or may be programmed to operate as a counter. *See* DIGITAL COMPUTER.

Counter specifications. Counters have progressed from relays to light-wavelength-geometry very large-scale-integrated circuits. There are several technologies for building individual digital counters. Single counters are available as integrated circuit chips in emitter-coupled logic (ECL), transistor-transistor logic (TTL), and CMOS. The three technologies are listed in the order of decreasing speed and decreasing power dissipation. Commercially available integrated circuits will operate at frequencies up to 4 GHz, with experimental Josephson junction counters capable of 10 times faster operation. *See* LOGIC CIRCUITS; SUPERCONDUCTING DEVICES.

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Digital evidence

Text, numerals, sound, or images (both moving and still) that are either stored or transmitted in a binary form and have potential use as evidence to prove a circumstance or event in a court of law. Binary refers to the electronic code made up of ones and zeroes that defines all the digital information. A unit of information is assigned one of two states, either 1 or 0 (that is, on and off). Thes two digits used in groups of 8 with varying sequences, for example, define the alphabet. *See* CRIMINALISTICS.

At a crime scene, investigators look for clues and motives among many items, including papers and documents such as daily appointment calendars, address books, letters, business records, checkbooks, and notes. Today, investigators are likely to find this information recorded on electronic media. Just like

paper records, digital information found on computers, personal organizers, answering machines, and cellular phones represents potentially latent (hidden) evidence that is important in investigations. For example, latent fingerprints are made visible after chemically treating surfaces. Likewise, latent evidence on computers and other digital devices is identified using computer hardware and software. Digital evidence on a cell phone may include an address book with names and telephone numbers, as well as a log of calls received, missed, made, and the times of these calls. On a personal computer, the log of Internet sites visited and e-mail sent and received might be important evidence in an investigation. *See* DATA COMMUNICATIONS; DIGITAL COMPUTER; ELECTRONIC MAIL; INTERNET; SOFTWARE.

Evidence collection. Computer forensics has come to mean the identification, preservation, extraction, documentation, and interpretation of digital evidence on computers or other electronic media for evidentiary purposes.

Forensic scientists are trained to consider evidence from both a scientific and legal perspective. When a scene is processed for evidence, several legal principles must be followed. For instance, the potential evidence must be identified, collected, transported, examined, and stored in a manner that does not change it. At the time the evidence is collected, it must be marked to identify who collected it, the date and time of collection, and where it was collected to preserve the chain of custody for acceptance in a court of law.

The law requires that the "best evidence" be used for legal purposes. In the case of a paper document, the original is preferred, although a copy can be used if the original is not available. Exact copies of digital evidence can be made from the original evidence. At a crime scene, the data on hard drives, disks, or other electronic media are copied and examined to find any evidence they might contain. The Scientific Working Group on Digital Evidence (SWGDE) has drafted the following definitions to describe the various attributes of digital evidence. *See* COMPUTER STORAGE TECHNOLOGY.

Acquisition of digital evidence: the process that begins when information or physical items are collected or stored for examination purposes. Applying the term "evidence" to these items implies that the manner in which they were collected is recognized by the courts; that is, the collection process was appropriate according to the rules of evidence in a given locality. A data object or physical item becomes evidence only when so deemed by a law enforcement official or designee.

Data objects: objects or information associated with physical items. Data objects may occur in different formats without altering the original information. Data objects include logical files, slack (unused) space, or perhaps a physical sector located on the physical item.

Physical item: an item on which data objects or information is stored or through which data objects are transferred.

Original digital evidence: physical items and the data objects associated with such items at the time of acquisition or seizure.

Duplicate digital evidence: an accurate digital reproduction of all data objects contained on an original physical item.

Evidence examination. Generally, the copy of a suspect's hard drive taken at the crime scene is the best evidence, and is stored in the evidence room as "the copy." Duplicates are made from this copy for forensic scientists and law enforcement to work from, with all work done on a dedicated forensic computer that is not connected to the Internet. This computer workstation is dedicated for forensic examination and is used in a controlled environment. The task for the forensic examiner is to locate documents or phrases that will assist the criminal investigator. For example, if a bomb was sent through the mail using a computer-printed address label, the copy of the suspect's hard drive would be examined for evidence of the address. In this case, the forensic examiner would do a keyword search and look for the words in the address. Forensic tools are software specifically designed for forensic examinations such as imaging (copying the hard drive) and restoring deleted files. Comprehensive suites of software tools are used for most forensic examinations.

Hidden data. Digital evidence may be obvious to an investigator or hidden. For example, obvious evidence might be a letter stored as a document on a hard drive, while less obvious evidence might be a word or fragment of a word that was deleted from a document but still exists on the unused space of the hard drive. Special software tools, some public and others restricted, assist forensic examiners to find this evidence. There are also forensic tools for determining when data have been deliberately hidden.

Encryption. Encryption is one way of actively hiding data. In encryption, data are encoded using an algorithm (mathematical formula), obscuring it from anyone who does not have a "secret key" to unscramble it. Once data have been encrypted, a single- or double-key decryption program is needed to read it, or else the code must be "cracked." Single-key decryption is similar to using one key to lock and unlock your home. Double-key decryption requires both a public key and a private key, where one or more people may have the public key, but only one person has the private key. Code cracking ranges from a very easy process to one that is computationally infeasible depending on the sophistication of the encryption. *See* ALGORITHM; CRYPTOGRAPHY.

Steganography. Hiding data in text, image, or audio files is called steganography. For example, a text file can be hidden in a digital picture. There are statistical analysis techniques to find such data, but they take time and effort. Commercial tools are available to detect these types of files.

Evidence authentication. A mathematical method known as hashing is used to verify the authenticity of data copied from a hard drive, disk, or file. A hash is a small mathematical summary of a data file or message. Changing one character in a file will change

the numerical value of the hash. If there has been any change in the file during transmission or from tampering, the hash number will change. A hash of the hard drive is done before copying it. The hash value of a hard drive and the copy should be exactly the same number.

To ensure chain of custody, forensic scientists must sign their name to the evidence. In the case of digital evidence, an authentic digital signature or a regular physical signature is required. Creating a digital signature is a simple process for verifying the integrity and originator of a message, whereby the signer calculates the hash value of the message, and then encrypts the hash value with his or her own private key. The resultant message "digest" is the digital signature. Appending it to the original message qualifies the object as being digitally signed.

As technology changes, the nature and handling of digital evidence will remain a challenge for forensic scientists. The wireless Web will allow for "virtual evidence rooms" where evidence is stored and secured through encryption. The chain of custody will be maintained through digital signatures, and the evidence will be checked out of the virtual evidence room with a time stamp from a digital clock synchronized with the atomic clock run by the National Institute of Standards and Technology (NIST).

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Digital filter

Any digital computing means that accepts as its input a set of one or more digital signals from which it generates as its output a second set of digital signals. While being strictly correct, this definition is too broad to be of any practical use, but it does demonstrate the possible extent of application of digital-filter concepts and terminology.

Capabilities. Digital filters can be used in any signal-manipulating application where analog or continuous filters can be used. The classic applications are low-pass, high-pass, band-pass, band-stop, and narrow-band notch filters. Classic filter forms are Butterworth, Chebyshev, inverse Chebyshev, elliptic, and Bessel. Digital filters can provide these response characteristics and much more. They can be used for such applications as emphasis or attenuation of specified frequencies for spectrum shaping and equalization, for time-domain pulse shaping in data transmission, for signal detection by matched filtering, and for noise or interference removal. Because of their utterly predictable performance, they can be used in exacting applications where analog filters fail because of time- or other parameter-dependent coefficient drift in continuous systems. Because of the ease and precision of setting the filter coefficients, adaptive and learning digital filters are comparatively simple and particularly effective to implement. In many applications, digital filters provide the least costly solutions. As digital technology becomes more ubiquitous, digital filters are increasingly acknowledged as the most versatile and cost-effective solutions to filtering problems.

The number of functions that can be performed by a digital filter far exceeds that which can be performed by an analog, or continuous, filter. By controlling the accuracy of the calculations within the filter (that is, the arithmetic word length), it is possible to produce filters whose performance comes arbitrarily close to the performance expected of the perfect models. For example, theoretical designs that require perfect cancellation can be implemented with great fidelity by digital filters.

Although the independent variable of the input to the digital filter could be almost any parameter, for the sake of clarity and simplicity time will be chosen as the independent variable, unless otherwise specifically stated.

Digital signals. A digital signal, in the sense used here, is a signal that has been sampled or quantized in time (that is, it is a discrete-time signal, quantized in amplitude, and finally encoded, so that its value can be expressed in a finite set of characters, such as binary words or symbols (such as numbers) composed of bits (1's and 0's). Examples of signals are the set of temperature measurements taken at specified times in a chemical-processing vat, the outputs from navigation sensors (angle, velocity, or position as a function of time), the sensed microvoltage on the antenna in a communications system (which contains the voice message to be extracted), the operational engine parameters (temperatures, pressures, and flow rates as functions of time) in the space shuttle, and the video intermediate-frequency voltage in a television receiver. *See* ANALOG-TO-DIGITAL CONVERTER; PULSE MODULATION.

Two more simplifying assumptions can be made. Although a digital filter could accept many inputs and generate many outputs simultaneously, this discussion is restricted to the most common case, the so-called single-input, single-output (SISO) case, as opposed to the multivariable case. Furthermore, although the time variable could be sampled in an infinite number of ways, the sampling is assumed to be uniform (which is not necessary, but is the usual and simplest case), so that while $x(t)$ is a function of the continuous variable, t, only the equally spaced sample values $x(nT) = x_n$ are considered. The sampling period, $T = 1/f_s$, is fixed. The sampling frequency is f_s and is assumed to be constant; therefore, the much-preferred and simpler subscripted form (that is, x_n) of the variables is unambiguous.

Linear difference equation. The digital filter can be a program run on a personal computer, a circuit card containing many electronic components,

an application-specific integrated circuit, or a programmable microprocessor device. Although many embodiments are possible, they all fit the following description. The digital filter accepts as its input signals numerical values called input samples and produces as its output signal numerical values called output samples. The string of input samples is sometimes referred to as the input vector, and the string of output samples as the output vector. Each output sample at any particular sampling instant is a weighted sum of present and past input samples, and past output samples. If the sequence of input samples is x_n, x_{n+1}, x_{n+2}, . . ., then the corresponding sequence of output samples would be $y_n, y_{n+1}, y_{n+2}, \ldots$. The linear digital filter is uniquely characterized by its difference equation (1), which

$$y_n = a_0x_n + a_1x_{n-1} + a_2x_{n-2} + \cdots + a_Mx_{n-M} + b_1y_{n-1} + b_2y_{n-2} + \cdots + b_Ny_{n-N} \quad (1)$$

tells exactly how to determine the value of each output sample. *See* DIFFERENCE EQUATION.

From this simple time-domain expression, a considerable number of definitions can be constructed. If the filter coefficients (the a's and the b's) are independent of the x's and y's, this digital filter is a linear filter. If the a's and b's are fixed, this is a linear time-invariant (LTI) filter. The order of the filter is given by the largest of the subscripts among the a's and b's, that is, the larger of M and N. If the b's are all zero (that is, if the output is the weighted sum of present and past input samples only), the digital filter is referred to as a nonrecursive (having no feedback) or finite impulse response (FIR) filter because the response of the filter to an impulse (actually a unit pulse) input is simply the sequence of the a coefficients. If any value of b is nonzero, the filter is recursive (having feedback) and is generally (although there are pathological exceptions) an infinite impulse response (IIR) filter. Sometimes the set of a coefficients is considered as the input- (sample-) weighting coefficient vector, and the set of b coefficients is considered as the output- (sample-) weighting-coefficient vector. Alternatively, the a and b coefficients can be strung together to form one large coefficient vector.

The difference equation of any linear, time-invariant, single-input, single-output digital filter looks the same. In order to design such a digital filter, all that remains is to determine the numerical values of the coefficient vector. The design, of course, flows from the functional requirements.

Although the time-domain difference equation is a useful description of a filter, as in the continuous-domain filter case, a powerful alternative form is the transfer function. The information content of the transfer function is the same as that of the difference equation as long as a linear, time-invariant system is under consideration. If the digital filter under consideration is not a linear, time-invariant filter, the transfer function cannot be used.

Transfer functions. A difference equation is converted to transfer-function form by use of the z transform. The z transform is simply the Laplace transform adapted for sampled systems with some shorthand notation introduced. The only operations on signals that are available in continuous, linear, time-invariant systems are addition (or subtraction), scaling (multiplication by a constant), and differentiation (or integration). Storage, or memory, in continuous systems is embodied through integration. The Laplace variable is s, and so is the Laplace operator for differentiation; the Laplace operator for integration is $1/s$. Similarly, the only allowable operations on signals in discrete-time, linear, time-invariant systems are addition (or subtraction), scaling (multiplication by a constant), and delay. Storage, or memory, in discrete-time systems is implemented directly. The Laplace operator for T-second storage is $\exp(-sT)$. *See* LAPLACE TRANSFORM; Z TRANSFORM.

The discrete-time version of the Laplace variable is simply the Laplace operator for storage. Because the notation $\exp(-sT)$ is so unwieldy, the shorthand time-shift variable $z = \exp(sT)$ is used, and has become a complex variable of interest in its own right. There is then an analogy: The continuous-time memory (storage) operator is $1/s$; the discrete-time memory (storage) operator is $1/z$. The z transform of x_n is $X(z)$; the z transform of x_n, scaled by α and delayed p samples (αx_{n-p}), is $\alpha X(z)z^{-p}$. The z transform of Eq. (1) is Eq. (2). The terms in this equation can be

$$Y(z) = \sum_{m=0}^{M} a_m \cdot X(z) \cdot z^{-m} + \sum_{n=1}^{N} b_n \cdot Y(z) \cdot z^{-n} = X(z) \cdot \sum_{m=0}^{M} a_m \cdot z^{-m} + Y(z) \cdot \sum_{n=1}^{N} b_n \cdot z^{-n} \quad (2)$$

rearranged to yield Eq. (3), and from this follows an expression for a rational polynomial transfer function, Eq. (4).

$$Y(z) \cdot \left[1 - \sum_{n=1}^{N} b_n \cdot z^{-n}\right] = X(z) \cdot \sum_{m=0}^{M} a_m \cdot z^{m} \quad (3)$$

$$\frac{Y(z)}{X(z)} = \frac{\sum_{m=0}^{M} a_m \cdot z^{m}}{1 - \sum_{n=1}^{N} b_n \cdot z^{-n}} \quad (4)$$

Since $z = \exp(sT)$, the complex (amplitude and phase) frequency response of the filter can be evaluated by setting $s = j\omega$, or $z = \exp(j\omega T)$. The design of the transfer function to meet a frequency-response specification is a well-established engineering procedure. *See* CONTROL SYSTEMS; LINEAR SYSTEM ANALYSIS.

FIR-filter design procedures. If $N = 0$ in Eqs. (1)–(4), an FIR filter results. The most common FIR-filter design approaches in computing the values of a_m are direct design using a procedure called the Remez-exchange algorithm (or the McClellan-Parks design procedure) and a Fourier-series design. Implementation of the filter is generally mechanized by either the equivalent of a tapped delay line (**Fig. 1***a*) or an injected delay line (Fig. 1*b*). The specific implementation choice is usually determined by characteristics

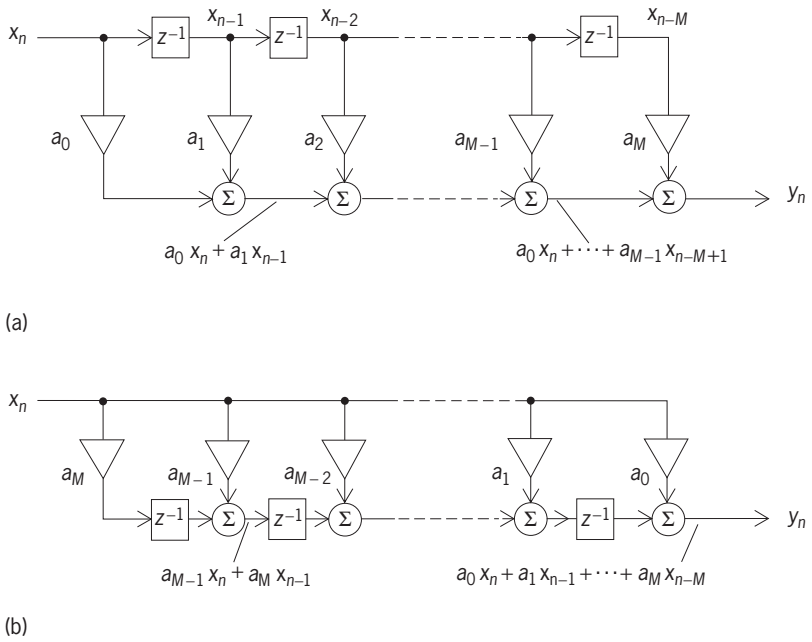

Fig. 1. Finite impulse response (FIR) filters. (*a*) Tapped delay-line filter. (*b*) Injected delay-line filter.

of the hardware used to build the filter. The delay-line approaches are replaced by fast convolution when the filter lengths become very long. Fast convolution consists of taking the Fourier transform of the input signal, multiplying the frequency-domain values by the Fourier transform of the desired filter response, then taking the inverse Fourier transform of the product. Fast convolution is very attractive and can have a mechanization-efficiency advantage over the delay-line approaches for FIR-filter lengths as short as $M = 15$. *See* FOURIER SERIES AND TRANSFORMS.

IIR-filter design procedures. For an IIR filter, a direct design approach such as one called the Deczky method can be used to determine the a_m and b_n values. Alternatively, one of the numerous cataloged continuous-time designs that have resulted from decades of work on continuous-time filters can be digitized by using one of the well-established transformation algorithms such as the bilinear-z transformation (usually the best), the matched-z transformation, or a transient-invariant z transformation (often better for certain control-systems applications). It should be emphasized that these are not z transforms; they are transformations.

Although obtaining the transfer function of the filter is (generally) unambiguously simple, the best mechanization of the IIR-filter transfer function is not correspondingly simple because the filters are mechanized with finite-resolution arithmetic, giving rise to three potential problems. The first problem is that the pole and zero locations are slightly in error because the coefficients of the transfer functions are inexact approximations to usually transcendental numbers. The second problem is that the multiplication and addition are performed with finite-precision accuracy, causing round-off quantization noise. The third problem is that the filter and signal must be scaled so that the amplitude of the signal can be adequately accommodated over the expected maximum and minimum values at all points within the filter.

These problems can be minimized by mechanizing the filter with the smallest-order sections possible. The numerator and denominator of the transfer function have real coefficients; therefore, each may

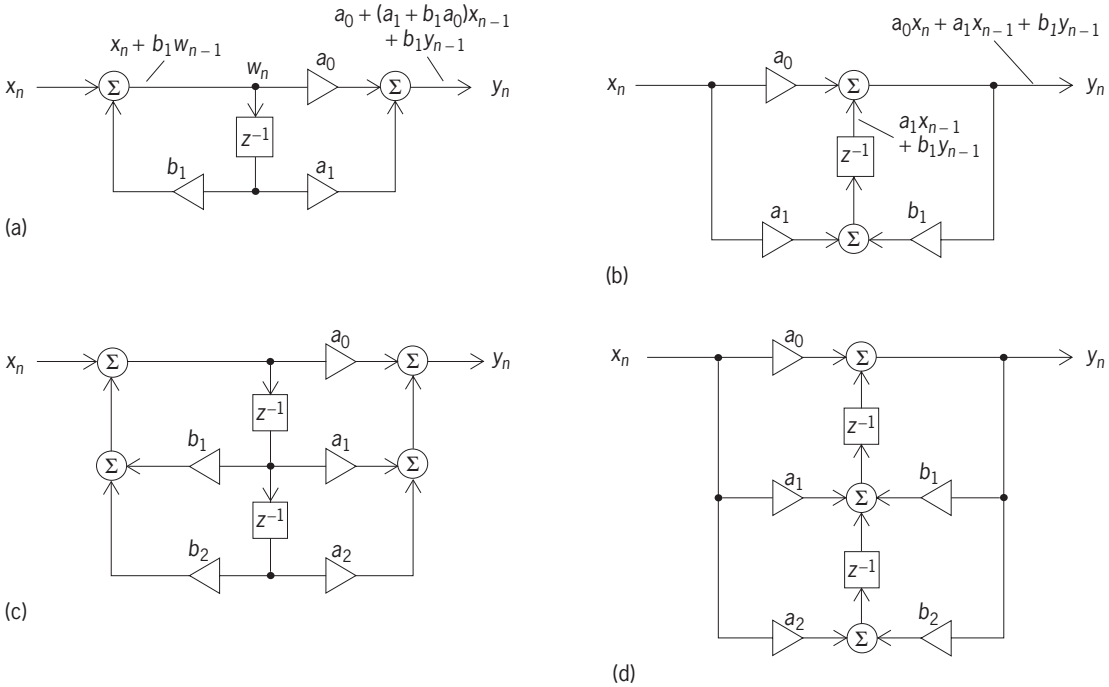

Fig. 2. Elemental building blocks of infinite impulse response (IIR) filters. (*a*) First-order building block; direct form. $w_n = b_1 w_{n-1} + x_n$, $y_n = a_0 w_n + a_1 w_{n-1}$. (*b*) First-order building block; transpose form. (*c*) Biquadratic building block; direct form. (*d*) Biquadratic building block; transpose form.

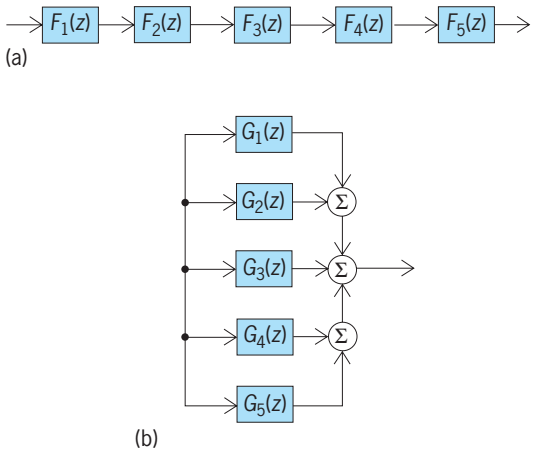

Fig. 3. Realizations of higher-order filters from elemental building blocks. (*a*) Cascade realization. (*b*) Partial-fraction-expansion realization.

be decomposed into a product of first-order factors with real roots (a numerator root is called a zero, a denominator root is called a pole) and second-order factors with complex-conjugate roots. The polynomial numerator and factored denominator may be further decomposed into alternate forms: a partial-fraction expansion, a cascade of first- and second-order building blocks, a parallel expansion of all-pass filters, a tapped cascade of all-pass filters, and the related families of lattice, ladder, and wave structures. Almost all practical linear, time-invariant, IIR digital filters are made of first- and second-order building blocks. *See* INTEGRATION.

Both first-order pole and zero IIR digital-filter building blocks and biquadratic (biquad) second-order building blocks may be realized in either direct or transpose configurations (**Fig. 2**). An intermediate state variable, w, is used as the storage variable in the direct form of the first-order building block (Fig. 2*a*). The building blocks in the partial-fraction realization of a high-order digital filter are similar to the first- and second-order building blocks in the cascade realizations (**Fig. 3**): the poles are the same as in the cascade realization, but the zeros are of order one less than the poles in each block.

All-pass filter sections are frequently used as building blocks (**Fig. 4**) because they have low coefficient sensitivity and are remarkably easy to scale. Their first- and second-order transfer functions are given by Eqs. (5). The magnitude-frequency responses of

$$H_1(z) = \frac{z^{-1} - b_1}{1 - b_1 \cdot z^{-1}}$$
$$H_2(z) = \frac{z^{-2} - b_1 \cdot z^{-1} + b_2}{1 - b_1 \cdot z^{-1} + b_2 \cdot z^{-2}} \tag{5}$$

all-pass filters are unity everywhere. These filter sections control phase alone.

Adaptive filters. So far only LTI filters have been discussed. An important class of variable-parameter filters change their coefficients to minimize an error criterion. These filters are called adaptive because they adapt their parameters in response to changes in the operating environment. An example is the FIR digital filter given by Eq. (6), where the a_m

$$y_n = \sum_{m=0}^{M} a_m \cdot x_{n-m} \tag{6}$$

are continually adjusted so that the output, y_n, will track a reference signal, r_n, with minimum error, ϵ. The performance criterion will be the minimization of some function of the error; that is, a function $J(\epsilon)$ will be minimized, where the rate of adjustment of each filter parameter is proportional to its contribution to $J(\epsilon)$, as in Eq. (7). This ad-

$$\frac{da_m}{dt} = -\mu \frac{\partial J(\varepsilon)}{\partial a_m} \tag{7}$$

justment method is known as the steepest-descent parameter-adjustment method. The convergence parameter is μ. If the convergence parameter is too

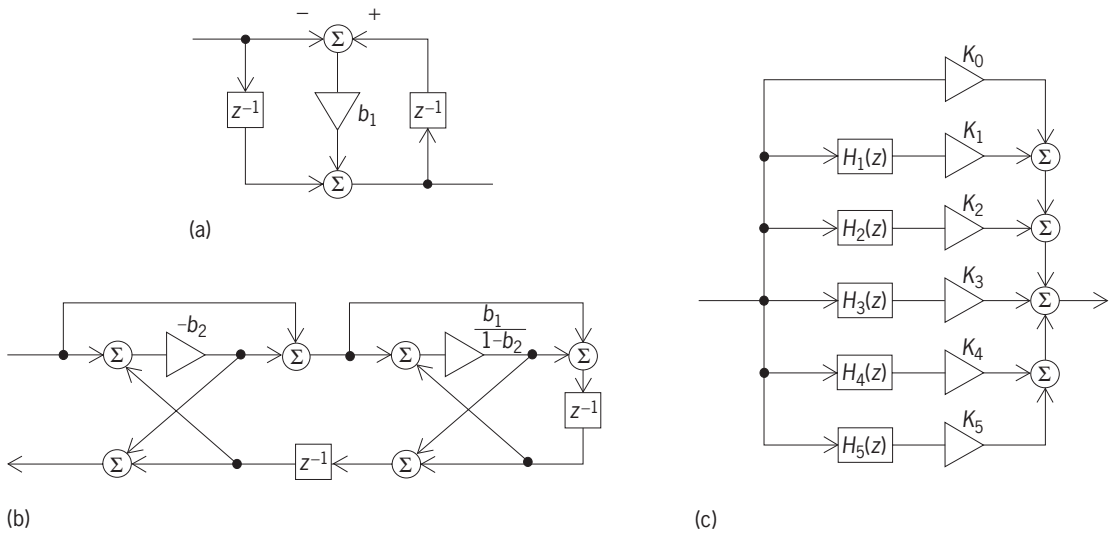

Fig. 4. Digital all-pass filters. (*a*) First-order filter. (*b*) Second-order lattice filter. (*c*) Higher-order filter assembled from a parallel bank of all-pass sections.

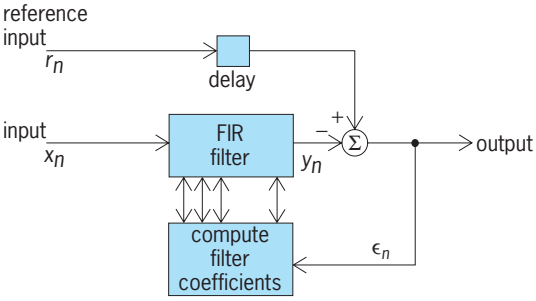

Fig. 5. Typical adaptive finite impulse response (FIR) filter.

small, convergence is very slow. If the convergence parameter is too large, the solution will oscillate around the correct value, never settling down. The error functional, $J(\epsilon)$, is generally of the form of Eq. (8), where κ is an integer. If $\kappa = 1$, the crite-

$$J(\varepsilon) = |\varepsilon|^{\kappa} \tag{8}$$

rion is least-absolute value; if $\kappa = 2$, the criterion is minimum-squared error (which is an unbiased estimate of the minimum mean-squared error, or MMSE, solution); and so forth (**Fig. 5**).

If the reference input to the filter (r_n) is signal plus additive noise and the "regular" input (x_n) is a linear-filtered version of the noise, then the adaptive filter will attempt to remove the noise from the output. This method is used in active noise-canceling systems. If both inputs are tied together, the filter will try to whiten the output; that is, the filter will attempt to remove correlated components such as echos and steady-state tones. By changing the effective time-constant of the adaption loops (making the integrators that generate the filter coefficients into low-pass filters by adding a low-gain "forgetting" loop), the correlation time is changed. Accordingly, the filter will emphasize (or deemphasize) signals with different correlation times, providing a powerful separation tool for signals such as speech. *See* ACTIVE SOUND CONTROL; SPEECH RECOGNITION.

There are also digital recursive adaptive filters (RAFs), which are used for more complicated problems. Although the mathematics is more involved because the error-criterion surface is multimodal (lumpy, rather than smooth), the basic idea is the same. *See* ADAPTIVE CONTROL; ELECTRIC FILTER; EQUALIZER.

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Digital-to-analog converter

A device for converting information in the form of combinations of discrete (usually binary) states or a signal to information in the form of the value or magnitude of some characteristics of a signal, in relation to a standard or reference. Most often, it is a device which has electrical inputs representing a parallel binary number, and an output in the form of voltage or current.

Figure 1 shows the structure of a typical digital-to-analog converter. The essential elements, found even in the simplest devices, are enclosed within the dashed rectangle. The digital inputs, labeled u_i, $i = 1, 2, \ldots, n$, are equal to 1 or 0. The output voltage E_o is given by Eq. (1), where V_{REF} is an analog reference

$$E_o = KV_{\text{REF}} \frac{u_1 2^{n-1} + u_2 2^{n-2} + \cdots + u_n 2^0}{2^n} \tag{1}$$

voltage and K is a constant. Thus, for a 5-bit binary converter with latched input code 10110, the output is given by Eq. (2).

$$E_o = \left(\frac{16}{32} + \frac{0}{32} + \frac{4}{32} + \frac{2}{32} + \frac{0}{32}\right) KV_{\text{REF}} = \frac{11}{16} KV_{\text{REF}} \tag{2}$$

Bit 1 is the "most significant bit" (MSB), with a weight of $\frac{1}{2}$; bit n is the "least significant bit" (LSB), with a weight of 2^{-n}. The number of bits n characterizes the resolution. [When the converter is used with a microprocessor, the bits are usually numbered from $n - 1$ (MSB) to 0 (LSB), that is, from 4 to 0 in the example of Eq. (2).] *See* ANALOG-TO-DIGITAL CONVERTER; NUMBERING SYSTEMS.

Uses. Digital-to-analog (D/A) converters (sometimes called DACs) are used to present the results of digital computation, storage, or transmission, typically for graphical display or for the control of devices that operate with continuously varying quantities. D/A converter circuits are also used in the design of analog-to-digital converters that employ feedback techniques, such as successive-approximation and counter-comparator types. In such applications, the D/A converter may not necessarily appear as a separately identifiable entity.

Circuitry. The fundamental circuit of most D/A converters involves a voltage or current reference; a resistive "ladder network" that derives weighted currents or voltages, usually as discrete fractions of the reference; and a set of switches, operated by the digital input, that determines which currents or voltages will be summed to constitute the output.

Input and output of converter in Fig. 2

Digital input code			Analog output	
u_1	u_2	u_3	$-E_o$, volts	$\frac{E_o}{V_{\text{REF}}}$
0	0	0	0	0
0	0	1	1.25	1/8
0	1	0	2.5	2/8
0	1	1	3.75	3/8
1	0	0	5.0	4/8
1	0	1	6.25	5/8
1	1	0	7.5	6/8
1	1	1	8.75	7/8

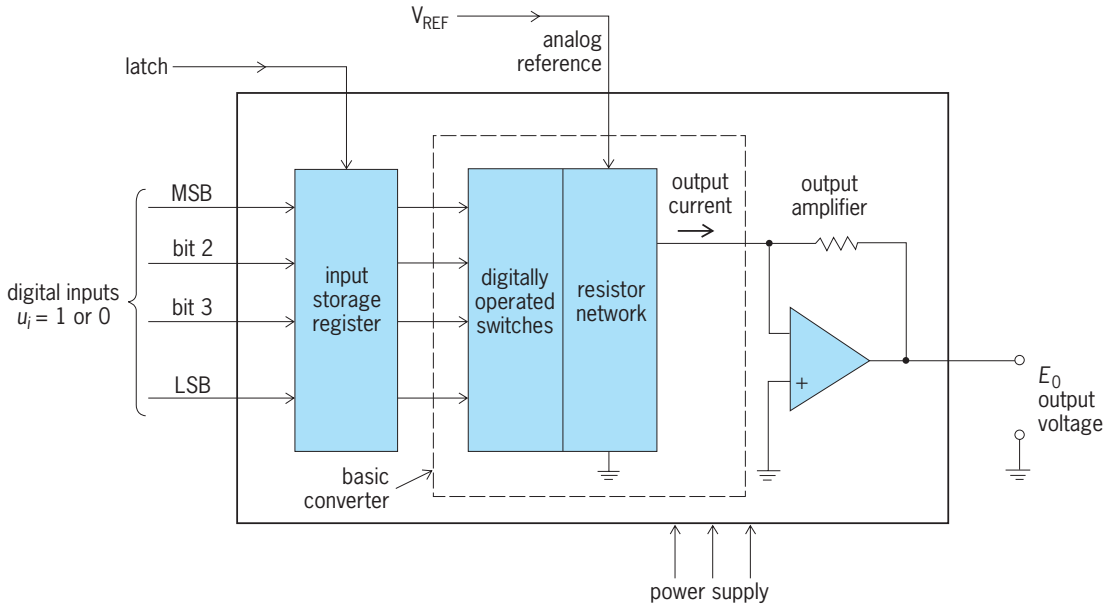

Fig. 1. Structure of a typical digital-to-analog converter.

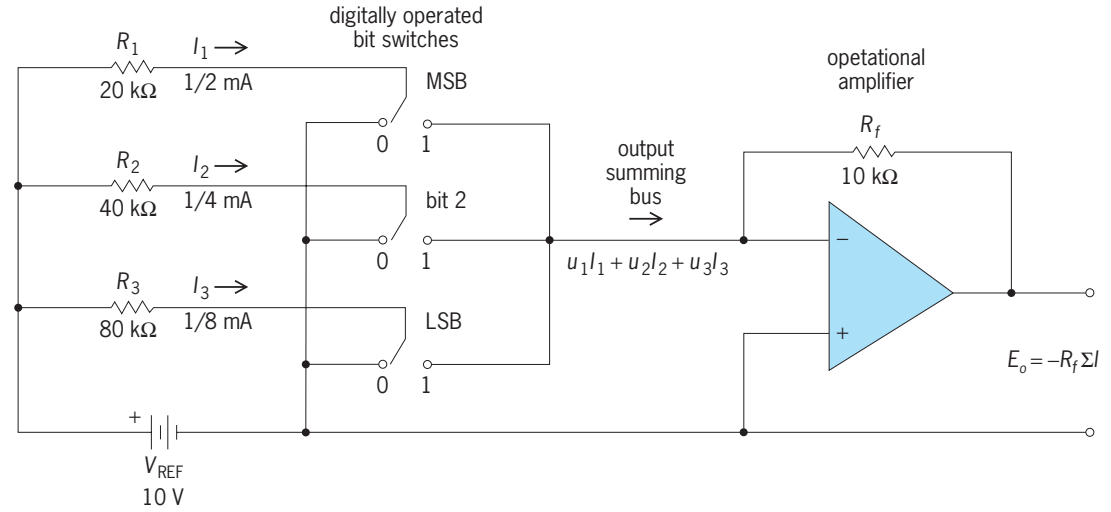

Fig. 2. Circuit of elementary 3-bit digital-to-analog converter.

An elementary 3-bit D/A converter is shown in **Fig. 2**. Binary-weighted currents developed in R_1, R_2, R_3 by V_{REF} are switched either directly to ground or to the output summing bus (which is held at zero volts by the operational-amplifier circuit). The sum of the currents develops an output voltage of polarity opposite to that of the reference across the feedback resistor R_f. The **table** shows the binary relationship between the input code and the output, both as a voltage and as a fraction of the reference.

The output of the D/A converter is proportional to the product of the digital input value and the reference. In many applications, the reference is fixed, and the output bears a fixed proportion to the digital input. In other applications, the reference, as well as the digital input, can vary; a D/A converter that is used in these applications is thus called a multiplying D/A converter. It is principally used for imparting a digitally controlled scale factor, or "gain," to an analog input signal applied at the reference terminal. *See* AMPLIFIER; ANALOG COMPUTER.

Construction. Except for the highest resolutions (beyond 18 bits), commercially available D/A converters are generally manufactured in the form of integrated circuits, using bipolar, MOS, and hybrid technologies. A single chip may include just the resistor network and switches; it may also include a reference circuit, output amplifier, and one or more sets of registers (with control logic suitable for direct microprocessor interfacing). *See* INTEGRATED CIRCUITS.

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Digitalis

A genus of the figwort family (Scrophulariaceae) ranging from the Canary Islands to central Asia. Foxglove (*Digitalis purpurea*), a native of western Europe, is the source of the important drug digitalis, much used in the treatment of heart disorders. Fresh, mature leaves are carefully selected, quickly dried, and stored in airtight containers. The active ingredient of digitalis is the glucoside digitalin. This slows and regulates the heartbeat, improving the tone and rhythm, and making the contractions more effective. The plant is also prized as a herbaceous perennial. *See* SCROPHULARIALES.

Perry D. Strausbaugh; Earl L. Core

Dill

An annual culinary herb of the carrot family, Apiaceae (Umbelliferae). It supplies four products: dill seed oil, dill leaf oil, dill seed, and dill leaf. Although all of the dill grown in the United States is one species, *Anethum graveolens,* a slightly different species, *A. sowa,* is grown in Asia. In the United States, two cultivars are grown, Mammoth and Bouquet. Bouquet is a tetraploid cultivar that was developed by treating Mammoth seed with colchicine and then selecting for a shorter (2–5 ft or 0.6–1.5 m) and more compact growth habit. Mammoth is a taller, more robust plant growing 4–8 ft (1.2–2.4 m) tall. Other varieties exist, such as those developed in India and Russia, but are not extensively grown here. Dill can be grown in most soil types in all areas of the United States, but most of the production is centered in Oregon, Washington, Florida, and California.

Flowering of dill usually occurs 60–80 days after planting, with hermaphrodite and later homogamous flowers being produced. Pollination is effected by insects—primarily bees and flies—but many other flying insects, including wasps, aid in fertilization. *See* POLLINATION.

Dill seed oil is a pale yellow oil which is steam-extracted just prior to seed shatter, about 105 to 120 days after planting. Dill grown for seed oil production is either planted in rows and furrow-irrigated or flat-planted and sprinkle-irrigated. Seeding rates vary from 4 to 20 lb/acre (3.5 to 18 kg/hectare). Most of this production provides oil used in the pickle industry for flavoring dill pickles. Dill leaf oil is little used, and relatively little dill is grown for this purpose. Dill seed is used as a spice, and large quantities of it are imported into the United States. India is the primary producer of dill seed for culinary use. *See* FAT AND OIL (FOOD).

Dill grown for leaf, often called dill weed, is cultivated in a slightly different manner. Seeding rates are 18–22 lb/acre (16–19 kg/ha), and it is planted in the fall or early spring to retard flowering. Dill leaf is harvested 50–70 days after planting. In the United States, dehydration is accomplished with forced-air dryers.

Weed control in commercial plantings is accomplished by use of herbicides, such as Treflan or Prometryn. Stoddards solvent or "carrot oil" at 40–60 gal/acre (380–570 liters/ha) has been extensively used in the past and is considered safe. *See* APIALES; SPICE AND FLAVORING.

Seth Kirby

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Dilleniales

An order of flowering plants, division Magnoliophyta (Angiospermae), subclass Dilleniidae, class Magnoliopsida (dicotyledons). It consists of 21 families: the Dilleniaceae, with about 350 species; and the Paeoniaceae, with a single genus of some 30 species. Within its subclass the order is marked by separate carpels; numerous stamens; bitegmic (having two integuments), crassinucellate (having a nucellus that is more than one cell thick) ovules; and seeds that have a well-developed endosperm and are provided with an aril (a partial or complete covering or appendage, usually derived from the funiculus of the ovule). The garden peony *(Paeonia lactiflora)* is a familiar member of the order (see **illus.**). *See* DILLENIIDAE; MAGNOLIOPHYTA; PLANT KINGDOM.

The garden peony (*Paeonia lactiflora*), of the family Paeoniaceae. (*F. E. Westlake, National Audubon Society*)

Arthur Cronquist; T. M. Barkley

Dilleniidae

A large subclass of the class Magnoliopsida (dicotyledons) of the division Magnoliophyta (Angiospermae), the flowering plants, consisting of 13 orders, 78 families, and nearly 25,000 species. The subclass is morphologically ill-defined, but most of the orders and species have the carpels united to form

a compound pistil. The petals are either separate or joined into a sympetalous corolla, but the sympetalous members generally lack the advanced features of the subclass Asteridae. That is, they do not have a single set of stamens alternate with the corolla lobes, ovules with a massive single integument, or a reduced nucellus. The stamens, when numerous, are initiated in centrifugal sequence. Many of the species have numerous ovules on parietal placentas; that is, the placentas are borne along the walls of an ovary which is usually with a single chamber. Many have various chemical repellents and are often tanniferous. There are mustard oils or iridoid compounds in some orders, but mostly they are poor in alkaloids and without betalains.

The largest orders included in the subclass are the Violales (about 5000 species), Capparales (about 4000 species), Malvales (about 3000 species), Theales (about 3500 species), and Ericales (about 4000 species). Other orders include the Dilleniales, Lecythidales, Nepenthales, Salicales, Diapensiales, Ebenales, Primulales, and Batales. See separate articles on each order. *See* ASTERIDAE; MAGNOLIOPHYTA; PLANT KINGDOM. Arthur Cronquist; T. M. Barkley

Dimensional analysis

A technique that involves the study of dimensions of physical quantities. Dimensional analysis is used primarily as a tool for obtaining information about physical systems too complicated for full mathematical solutions to be feasible. It enables one to predict the behavior of large systems from a study of small-scale models. It affords a convenient means of checking mathematical equations. Finally, dimensional formulas provide a useful cataloging system for physical quantities.

Theory. All the commonly used systems of units in physical science have the property that the number representing the magnitude of any quantity (other than purely numerical ratios) varies inversely with the size of the unit chosen. Thus, if the length of a given piece of land is 300 ft, its length in yards is 100. The ratio of the magnitude of 1 yd to the magnitude of 1 ft is the same as that of any length in feet to the same length in yards, that is, 3. The ratio of two different lengths measured in yards is the same as the ratio of the same two lengths measured in feet, inches, miles, or any other length units. This universal property of unit systems, often known as the absolute significance of relative magnitude, determines the structure of all dimensional formulas. *See* UNITS OF MEASUREMENT.

In defining a system of units for a branch of science such as mechanics or electricity, certain quantities are chosen as fundamental and others as secondary, or derived. The choice of the fundamental units is always arbitrary and is usually made on the basis of convenience in maintaining standards. In mechanics the fundamental units most often chosen are mass, length, and time. Standards of mass (the standard kilogram) and of length (the standard meter) are readily manufactured and preserved, while the rotation of the Earth gives a sufficiently reproducible standard of time. Secondary quantities such as velocity, force, and momentum are obtained from the primary set of quantities according to a definite set of rules.

Assume that there are three primary, or fundamental, quantities α, β, and γ (the following discussion, however, is not limited to there being exactly three fundamental quantities). Consider a particular secondary quantity, expressed in terms of the primaries as $F(\alpha,\beta,\gamma)$, where F represents some mathematical function. For example, if α = mass, β = length, and γ = time, the derived quantity velocity would be $F(\alpha,\beta,\gamma) = \beta/\gamma$.

Now, if it is assumed that the sizes of the units measuring α, β, and γ are changed in the proportions $1/x$, $1/y$, and $1/z$, respectively, then the numbers measuring the primary quantities become $x\alpha$, $y\beta$, and $z\gamma$, and the secondary quantity in question becomes $F(x\alpha, y\beta, z\gamma)$. Merely changing the sizes of the units must not change the rule for obtaining a particular secondary quantity.

Consider two separate values of the secondary quantity $F(\alpha_1,\beta_1,\gamma_1)$ and $F(\alpha_2,\beta_2,\gamma_2)$. Then, according to the principle of the absolute significance of relative magnitude, Eqs. (1*a*) and (1*b*) hold.

$$\frac{F(x\alpha_1, y\beta_1, z\gamma_1)}{F(x\alpha_2, y\beta_2, z\gamma_2)} = \frac{F(\alpha_1, \beta_1, \gamma_1)}{F(\alpha_2, \beta_2, \gamma_2)} \tag{1a}$$

$$F(x\alpha_1, y\beta_1, z\gamma_1) = F(x\alpha_2, y\beta_2, z\gamma_2)\frac{F(\alpha_1, \beta_1, \gamma_1)}{F(\alpha_2, \beta_2, \gamma_2)} \tag{1b}$$

Differentiating partially with respect to x, holding y and z constant, gives Eq. (2), where F' represents

$$\alpha_1 F'(x\alpha_1, y\beta_1, z\gamma_1) = \alpha_2 F'(x\alpha_2, y\beta_2, z\gamma_2)\frac{F(\alpha_1, \beta_1, \gamma_1)}{F(\alpha_2, \beta_2, \gamma_2)} \tag{2}$$

the total derivative of the function F with respect to its first argument.

Next set the coefficients $x, y, z = 1$. This gives Eq. (3). This relation must hold for all values of the

$$\frac{\alpha_1 F'(\alpha_1, \beta_1, \gamma_1)}{F(\alpha_1, \beta_1, \gamma_1)} = \frac{\alpha_2 F'(\alpha_2, \beta_2, \gamma_2)}{F(\alpha_2, \beta_2, \gamma_2)} \tag{3}$$

arguments α, β, and γ and hence is equal to a constant. The subscripts can now be dropped, giving Eq. (4). The general solution of this differential equation is $F = C_1\alpha^a$, where a is a constant and C_1 is in general a function of β and γ.

$$\frac{\alpha}{F}\frac{dF}{d\alpha} = \text{constant} \tag{4}$$

The above analysis can now be repeated for the parameters β and γ, leading to the results given in Eqs. (5*a*) and (5*b*). These solutions are consistent only if $F = C\alpha^a\beta^b\gamma^c$ where C, a, b, and c are constants. Thus every secondary quantity which satisfies

$$F = C_2(\alpha, \gamma)\beta^b \tag{5a}$$

$$F = C_3(\alpha, \beta)\gamma^c \tag{5b}$$

the condition of the absolute significance of relative magnitude is expressible as a product of powers of the primary quantities. Such an expression is known as the dimensional formula of the secondary quantity. There is no requirement that the exponents *a, b, c* be integral.

Examples of dimensional formulas. Table 1 gives the dimensional formulas of a number of mechanical quantities in terms of their mass M, length L, and time T.

In order to extend this list to include the dimensional formulas of quantities from other branches of physics, such as electricity and magnetism, one may do either of the following:

1. Obtain the dimensional formulas in terms of a particular unit system without introducing any new fundamental quantities. Thus, if one uses Coulomb's law in the centimeter-gram-second electrostatic system in empty space (dielectric constant = 1), one obtains definitions (6*a*) and (6*b*).

$$\text{Force} = \frac{(\text{charge})^2}{(\text{distance})^2} \tag{6a}$$

$$\text{Charge} = \text{distance} \times \text{square root of force} \tag{6b}$$

Thus, the dimensions of charge are $M^{1/2}L^{3/2}T^{-1}$. The dimensional formulas of other electrical quantities in the electrostatic system follow directly from definition (7) and so forth.

$$\begin{aligned} \text{Potential} &= \text{energy/charge} = M^{1/2}L^{1/2}T^{-1} \\ \text{Current} &= \text{charge/time} = M^{1/2}L^{3/2}T^{-2} \\ \text{Electric field} &= \text{force/charge} = M^{1/2}L^{-1/2}T^{-1} \end{aligned} \tag{7}$$

2. Introduce an additional fundamental quantity to take account of the fact that electricity and magnetism encompass phenomena not treated in mechanics. All electrical quantities can be defined in terms of *M, L, T,* and one other, without resorting to fractional exponents and without the artificial assumption of unit (and dimensionless) dielectric constant as in the electrostatic system. The usual choice for the fourth fundamental quantity is charge *A,* even though charge is not a preservable electrical standard. Then definitions (8) and so forth hold.

$$\begin{aligned} \text{Potential} &= \text{energy/charge} = ML^2T^{-2}Q^{-1} \\ \text{Current} &= \text{charge/time} = QT^{-1} \end{aligned} \tag{8}$$

See ELECTRICAL UNITS AND STANDARDS.

It must be realized that the choice of fundamental quantities is entirely arbitrary. For example, a system of units for mechanics has been proposed in which the velocity of light is taken as a dimensionless quantity, equal to unity in free space. All velocities are then dimensionless. All mechanical quantities can then be specified in terms of just two fundamental quantities, mass and time. This is analogous to the reduction of the number of fundamental quantities needed for electricity from four to three by taking the dielectric constant (permittivity) of empty space as unity. *See* DIMENSIONLESS GROUPS.

Furthermore, one could increase the number of fundamental quantities in mechanics from three to four by adding force F to the list and rewriting Newton's second law as $F = Kma$, where K is a constant of dimensions $M^{-1}L^{-1}T^2F$. One could then define a system of units for which K was not numerically equal to 1.

In the past there has been considerable controversy as to the absolute significance, if any, of dimensional formulas. The significant fact is that dimensional formulas always consist of products of powers of the fundamental quantities.

Applications. The important uses of the technique of dimensional analysis are considered in the following sections.

Checking of equations. It is intuitively obvious that only terms whose dimensions are the same can be equated. The equation 10 kg = 10 m/s, for example, makes no sense. A necessary condition for the correctness of any equation is that the two sides have the same dimensions. This is often a help in the verification of complicated analytic expressions. Of course, an equation can be correct dimensionally and still be wrong by a purely numerical factor.

A corollary of this is that one can add or subtract only quantities which have the same dimensions (except for the trivial case which arises when two different equations are added together). Furthermore, the arguments of trigonometric and logarithmic functions must be dimensionless; otherwise their power series expansions would involve sums of terms with different dimensions. There is no restriction on the multiplication and division of terms whose dimensions are different.

π Theorem. The application of dimensional analysis to the derivation of unknown relations depends upon the concept of completeness of equations. An expression which remains formally true no matter how the sizes of the fundamental units are changed

TABLE 1. Dimensional formulas of common quantities

Quantity	Symbol	Dimensional formula
Mass	Fundamental	M
Length	Fundamental	L
Time	Fundamental	T
Velocity	Distance/time	LT^{-1}
Acceleration	Velocity/time	LT^{-2}
Force	Mass × acceleration	MLT^{-2}
Momentum	Mass × velocity	MLT^{-1}
Energy	Force × distance	ML^2T^{-2}
Angle	Arc/radius	0
Angular velocity	Angle/time	T^{-1}
Angular acceleration	Angular velocity/time	T^{-2}
Torque	Force × lever arm	ML^2T^{-2}
Angular momentum	Momentum × lever arm	ML^2T^{-1}
Moment of inertia	Mass × radius squared	ML^2
Area	Length squared	L^2
Volume	Length cubed	L^3
Density	Mass/volume	ML^{-3}
Pressure	Force/area	$ML^{-1}T^{-2}$
Action	Energy × time	ML^2T^{-1}
Viscosity	Force per unit area per unit velocity gradient	$ML^{-1}T^{-1}$

is said to be complete. If changing the units makes the expression wrong, it is incomplete. For a body starting from rest and falling freely under gravity, $s = 16t^2$ is a correct expression only so long as the distance fallen s is measured in feet and the time t in seconds. If s is in meters and t in minutes, the equation is wrong. Thus, $s = 16t^2$ is an incomplete equation. The constant in the equation depends upon the units chosen. To make the expression complete, the numerical factor must be replaced by a dimensional constant, the acceleration of gravity g. Then $s = 1/2gt^2$ is valid no matter how the units of length and time are changed, since the numerical value of g can be changed accordingly.

Assume a group of n physical quantities $x_1, x_2, \ldots, x_n$, for which there exists one and only one complete mathematical expression connecting them, namely, $\phi(x_1, x_2, \ldots, x_n) = 0$. Some of the quantities $x_1, x_2, \ldots, x_n$ may be constants. Assume further that the dimensional formulas of the n quantities are expressed in terms of m fundamental quantities $\alpha, \beta, \gamma, \ldots$. Then it will always be found that this single relation ϕ can be expressed in terms of some arbitrary function F of $n - m$ independent dimensionless products $\pi_1, \pi_2, \ldots, \pi_{n-m}$, made up from among the n variables, as in Eq. (9). This is known as the π theorem.

$$F(\pi_1, \pi_2, \ldots, \pi_{n-m}) = 0 \tag{9}$$

It was first rigorously proved by E. Buckingham. The proof is straightforward but long. The only restriction on the π's is that they be independent (no one expressible as products of powers of the others). The number m of fundamental quantities chosen in a particular case is immaterial so long as they also are dimensionally independent. Increasing m by 1 always increases the number of dimensional constants (and hence n) by 1, leaving $n - m$ the same.

The main usefulness of the π theorem is in the deduction of the form of unknown relations. The successful application of the theorem to a particular problem requires a certain amount of shrewd guesswork as to which variables $x_1, x_2, \ldots, x_n$ are significant and which are not. If $\phi(x_1, x_2, \ldots, x_n)$ is not known, one can often still deduce the structure of $F(\pi_1, \pi_2, \ldots, \pi_{n-m}) = 0$ and so obtain useful information about the system in question.

An example is the swinging of a simple pendulum. Assume that the analytic expression for its period of vibration is unknown. Choose mass, length, and time as the fundamental quantities. Thus $m = 3$. One must make a list of all the parameters pertaining to the pendulum which might be significant. If the list is incomplete, no useful information will be obtained. If too many quantities are included, the derived information is less specific than it might otherwise be. The quantities given in **Table 2** would appear to be adequate.

The microscopic properties of the bob and string are not considered, and air resistance and the mass of the string are likewise neglected. These would be expected to have only a small effect on the period. Thus $n = 5$, $n - m = 2$, and therefore one expects to find

TABLE 2. Dimensional formulas for a simple pendulum

Quantity	Symbol	Dimensional formula
Mass of bob	m	M
Length of string	l	L
Acceleration of gravity	g	LT^{-2}
Period of swing	τ	T
Angular amplitude	θ	0

two independent dimensionless products. These can always be found by trial and error. There may be more than one possible set of independent π's. In this case one π is simply θ, the angular amplitude. Another is $l/\tau^2 g$. Since no other of the n variables contains M in its dimensional formula, the mass of the bob m cannot occur in any dimensionless product. The π theorem gives $F(\theta, l/\tau^2 g) = 0$. Therefore the period of vibration does not depend upon the mass of the bob.

Because τ appears in only one of the dimensionless products, this expression can be explicitly solved for τ to give $\tau = G(\theta)\sqrt{l/g}$, where $G(\theta)$ is an arbitrary function of θ. Now make the further assumption that θ is small enough to be neglected. Then $n = 4$, $n-m = 1$, and $F(l/\tau^2 g) = 0$. Thus $l/\tau^2 g =$ constant, or $\tau \propto \sqrt{l/g}$. The π theorem thus leads to the conclusion that τ varies directly as the square root of the length of the pendulum and inversely as the square root of the acceleration of gravity. The magnitude of the dimensionless constant (actually it is 2π) cannot be obtained from dimensional analysis.

In more complicated cases, where a direct solution is not feasible, this method can give information on how certain variables enter a particular problem even where $F(\pi_1, \pi_2, \ldots, \pi_{n-m}) = 0$ cannot be explicitly solved for one variable. The procedure is particularly useful in hydraulics and aeronautical engineering, where detailed solutions are often extremely complicated.

Model theory. A further application of dimensional analysis is in model design. Often the behavior of large complex systems can be deduced from studies of small-scale models at a great saving in cost. In the model each parameter is reduced in the same proportion relative to its value in the original system. *See* MODEL THEORY.

Once again the case of the simple pendulum is a good example. It was found from the π theorem that $F(\theta, l/\tau^2 g) = 0$. If the magnitudes of θ, l, τ, and g are now changed in such a way that neither argument of F is changed in numerical value, the system will behave exactly as the original system and is said to be physically similar. Evidently θ cannot be changed without altering any of the arguments of F, but l, g, and τ can be varied. Suppose that it was desired to build a very large and expensive pendulum which was to swing with finite amplitude. One could build a small model of, say, $^1/_{100}$ the length and time its swing for an amplitude equal to that for the desired pendulum. The acceleration of gravity g would be the same for the model as for the large pendulum.

The period for the model would then be just $^1/_{10}$ that for the large pendulum. Thus the period of the large pendulum could be deduced before the pendulum was ever built. In practice one would never bother with the π theorem in cases as simple as this where a full analytic solution is possible. In many situations where such a solution is not feasible, models are built and extensively studied before the full-scale device is constructed. This technique is standard in wind tunnel studies of aircraft design. *See* DYNAMIC SIMILARITY.

Cataloging of physical quantities. Dimensional formulas provide a convenient shorthand notation for representing the definitions of secondary quantities. These definitions depend upon the choice of primary quantities. The π theorem is applicable no matter what the choice of primary quantities is.

Changing units. Dimensional formulas are helpful in changing units from one system to another. For example, the acceleration of gravity in the centimeter-gram-second system of units is 980 cm/s^2. The dimensional formula for acceleration is LT^{-2}. To find find the magnitude of g in mi/h^2, one would proceed as in Eq. (10).

$$980\frac{\text{cm}}{s^2} \times \frac{(\text{conversion factor for length})}{(\text{conversion factor for time})^2}$$
$$= 980\frac{\text{cm}}{\text{s}^2} \times \frac{\dfrac{1}{1.609 \times 10^5}\dfrac{\text{mi}}{\text{cm}}}{(1/3600)^2(\text{h}^2/\text{s}^2)}$$
$$= 7.89 \times 10^5 \text{mi/h}^2 \qquad (10)$$

In the past the subject of dimensions has been quite controversial. For years unsuccessful attempts were made to find "ultimate rational quantities" in terms of which to express all dimensional formulas. It is now universally agreed that there is no one "absolute" set of dimensional formulas. Some systems are more symmetrical than others and for this reason are perhaps preferable. The representation of electrical quantities in terms of M, L, and T alone through the electrostatic form of Coulomb's law leads to somewhat awkward fractional exponents, but nevertheless is just as "correct" as a representation in which charge is used as a fourth fundamental unit.

A highly symmetrical pattern results if energy, linear displacement, and linear momentum are chosen as the fundamental quantities in mechanics. In electricity one can use energy, charge, and magnetic flux. The corresponding quantities for a vibrating mass on a spring and the analogous alternating-current circuit with inductance and capacitance have similar dimensional formulas. In this analogy energy is invariant, charge corresponds to displacement, and magnetic flux corresponds to linear momentum. This correspondence is not displayed in conventional dimensional formulas.

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Dimensionless groups

A dimensionless group is any combination of dimensional or dimensionless quantities possessing zero overall dimensions. Dimensionless groups are frequently encountered in engineering studies of complicated processes or as similarity criteria in model studies. A typical dimensionless group is the Reynolds number (a dynamic similarity criterion), $\text{Re} = vl\rho/\mu$. Since the dimensions of the quantities involved are velocity v: $[L/\theta]$; characteristic length l: $[L]$; density ρ: $[M/L^3]$; and viscosity μ: $[M/LT]$ (with M, L, and T as the fundamental units of mass, length, and time), the Reynolds number reduces to a dimensionless group and can be represented by a pure number in any coherent system of units. *See* DYNAMIC SIMILARITY.

History of named groups. The importance of dimensionless groups has long been known; as early as 1850 G. G. Stokes found the group now known as the Reynolds number and demonstrated its importance as a dynamic similarity criterion. In 1873 H. von Helmholtz showed the importance of the groups now known as the Froude and Mach numbers. The first attempt to attach names to important dimensionless groups appears to have been made in 1919 by M. G. Weber, who first named the Froude, Reynolds, and Cauchy numbers. It has become customary to name dimensionless groups after the outstanding or pioneer workers in the field. Groups are often denoted by the first two letters of the name; thus, the Reynolds number is usually denoted as Re.

Sources and uses. Many important problems in applied science and engineering are too complicated to permit completely theoretical solutions to be found. However, the number of interrelated variables involved can be reduced by carrying out a dimensional analysis to group the variables as dimensionless groups. *See* DIMENSIONAL ANALYSIS.

If almost nothing is known about the problem, it is necessary to decide which physical parameters will affect the process being studied. A dimensional analysis can then be carried out by the well-known π-theorem method of E. Buckingham or by the indicial method proposed by Lord Rayleigh. The success of these methods depends on identifying all of the important parameters for the problem. On the other hand, if the differential equations describing the phenomenon are known, inspectional analysis of the equations will yield all the important dimensionless groups even when the differential equations cannot be solved. Various procedures have been developed to determine the groups which correspond to the assumed parameters or variables. The form of the dimensionless equations linking these groups has to be determined by the use of experimental data in all except the simplest cases. Often there is no better procedure available in the case of complicated phenomena.

The advantages of using dimensionless groups in studying complicated phenomena include:

1. A significant reduction in the number of "variables" to be investigated; that is, each dimensionless

group, containing several physical variables, may be treated as a single compound "variable," thereby reducing the number of experiments needed as well as the time required to correlate and interpret the experimental data.

2. Predicting the effect of changing one of the individual variables in a process (which it may be impossible to vary much in available equipment) by determining the effect of varying the dimensionless group containing this parameter (this must be done with some caution, however).

3. Making the results independent of the scale of the system and of the system of units being used.

4. Simplifying the scaling-up or scaling-down of results obtained with models of systems by generalizing the conditions which must exist for similarity between a system and its model.

Deducing variation in importance of mechanisms in a process from the numerical values of the dimensionless groups involved; for instance, an increase in the Reynolds number in a flow process indicates that molecular (viscous) transfer mechanisms will be less important relative to transfer by bulk flow (inertia effects), since the Reynolds number is known to represent a measure of the ratio of inertia forces to viscous forces. *See* UNITS OF MEASUREMENT.

Classification. A large number of independent dimensionless groups may arise from a dimensional analysis of a complex problem; this number usually equals the number of variables in the problem minus the number of fundamental units in the system of units used. Furthermore, since any product or quotient of powers of dimensionless groups is also dimensionless, it follows that an almost limitless number of alternative dimensionless groups can be generated. Several hundred groups are known, many of them named (see **table**). It is therefore useful to be able to specify the conditions for obtaining a complete statement of a problem in terms of dimensionless groups and to classify the large number of groups in some way.

Complete sets. A set of dimensionless groups of given variables is complete if each group in the set is independent of the others and if every other dimensionless combination of the variables is a product of powers of these groups. For example, a complete set of dimensionless groups for a problem involving the variables length (l), velocity (v), density (ρ), viscosity (μ), the acceleration of gravity (g), the velocity of sound (v_s), and surface tension (γ) would be the Reynolds number, $\mathrm{Re} \equiv vl\rho/\mu$; the Froude number, $\mathrm{Fr} = v/(gl)^{1/2}$; the Mach number, $\mathrm{Ma} = v/v_s$, and the Weber number, $\mathrm{We} = v(\rho l/g)$. None of these groups is a product of powers of the others, since μ occurs only in Re, g only in Fr, v_s only in in Ma, and γ only in We. Many other

TABLE 1. List of frequently occurring dimensionless groups*

Symbol and name	Definition	Notation	Physical significance
Ar; Archimedes number	$d_p^3 g\rho_f(\rho_s-\rho_f)/\mu^2$		(Inertia force) (gravity force)/(Viscous force)^2
Bi; Biot number	hl/k_s	k_s = thermal conductivity of solid, $ML/T^3\theta$	(Internal thermal resistance)/(Surface thermal resistance)
Bi_m; mass transport Biot number	$k_m L/D_i$	L = Layer thickness, L D_i = interface diffusivity, L^2/T	(Mass transport conductivity at solid/fluid interface)/(Internal transport conductivity of solid wall of thickness L)
Bo; Bond number (also Eo, Eötvös number)	$d^2g(\rho-\rho_f)/\gamma$	d = bubble or droplet diameter, L	(Gravity force)/(Surface tension force)
Bq; Boussinesq number	$v/(2gm)^{1/2}$	m = mean hydraulic depth of open channel, L	(Inertia force)$^{1/2}$/(Gravity force)$^{1/2}$
Dn; Dean number	$(d/2R)^{1/2}(\mathrm{Re})$	d = pipe diameter, L R = radius of curvature of channel centerline, L Re = Reynolds number	Effect of centrifugal force on flow in a curved pipe
Ec; Eckert number	$v^2/C_p\Delta\theta$	$\Delta\theta$ = temperature difference, θ	Used in study of variable-temperature flow
Ek; Ekman number	$(\nu/2\omega l^2)^{1/2}$		(Viscous force)$^{1/2}$/(Coriolis force)$^{1/2}$
f; Fanning friction factor	$2\tau_w/\rho v^2 = D\,\Delta p/2L\rho v^2$	τ_w = wall shear stress, M/LT^2 D = pipe diameter, L L = pipe length, L	(Wall shear stress)/(Velocity head)
Fo_f; Fourier flow number	$\nu t/l^2$	—	Used in undimensionalization
Fo; Fourier flow number	$\alpha t/l^2$	—	Indicates the extent of thermal penetration in unsteady-state heat transport
Fo_m; mass transport Fourier number	$k_m t/l$	—	Indicates the extent of substance penetration in unstedy-state mass transport
Fr; Froude number	$v/(gl)^{1/2}$	—	(Inertia force)$^{1/2}$/(Gravity force)$^{1/2}$
Ga; Galileo number	l^3g/ν^2	—	(Inertia force)(gravity force)/(Viscous force)^2
Gz; Graetz number	$\dot{m}C_p/k_f l$	$\dot{m}$ = mass flow rate, M/T C_p = specific heat capacity (constant pressure), $L^2/T^2\theta$	(Fluid thermal capacity)/(Thermal energy transferred by conduction)
Gr; Grashof number	$l^3g\Delta\rho/\rho_f\nu^2$	$\Delta\rho$ = density driving force, M/L^3	(Inertia force)(buoyancy force)/(Viscous force)^2
Gr_m; mass transport Grashof number	$l^3g\beta_c\Delta c l\nu^2$	β_c = volumetric expansion coefficient Δc = concentration driving force	(Inertia force)(buoyancy force)/(Viscous force)^2
Ha; Hartmann number (M)	$lB(\sigma/\mu)^{1/2}$	B = maganetic flux density, M/QT σ = electrical conductivity, Q^2T/ML^3	(Magnetically induced stress)$^{1/2}$/(Viscous shear stress)$^{1/2}$
Kn; Knudsen number	λ/l	λ = length of mean free path	Used in study of low-pressure gas flow
Le; Lewis number	D/α	—	(Molecular diffusivity)/(Thermal diffusivity)
Lu; Luikov number	$k_m l/\alpha$	—	(Mass diffusivity)/(Thermal diffusivity)
Ly; Lykoudis number	—	—	(Hartman number)^2/(Grashof number)$^{1/2}$
Ma; Mach number	v/v_s	v_s = velocity of sound in fluid, L/T	(Linear velocity)/(Velocity of sound)

(*continued*)

TABLE 1. List of frequently occurring dimensionless groups* (*cont.*)

Symbol and name	Definition	Notation	Physical significance
Ne; Newton number	$F/\rho_f v^2/^2$	F = hydrodynamic drag force, ML/T^2	(Resistance force)/(Inertia force)
Nu; Nusselt number	hl/k_f	—	(Thermal energy transport in forced convection)/(Thermal energy transport if it occurred by conduction)
Pe; Peclet number	lv/α	—	(Reynolds number)(Prandtl number); (Bulk thermal energy transport in forced convection)/(Thermal energy transport by conduction)
Pe_m; mass transport Peclet number	lv/D	—	(Bulk mass transport)/(Diffusional mass transport)
Ps; Poiseuille number	$v\mu/gd_p^2(\rho_s \quad \rho f)$	—	(Viscous force)/(Gravity force)
Pr; Prandtl number	ν/α	—	(Momentum diffusivity)/(Thermal diffusivity)
Ra; Rayleigh number			(Grashof number) (Prandtl number)
Ra_m; mass transport Rayleigh number			(Mass transport Grashof number) (Schmidt number)
Re; Reynolds number	vl/ν	—	
Re_R; rotational Reynolds number	l^2N/ν	N = rate of rotation (revolution per time), $1/T$	(Inertia force)/(Viscous force)
Sc; Schmidt number	ν/D	—	(Momentum diffusivity)/(Molecular diffusivity)
Sh; Sherwood number	$k_m l/D$	—	(Mass diffusivity)/(Molecular diffusivity)
St; Stanton number	$h/C_p\rho v$	—	(Nusselt number)/(Reynolds number) (Prandtl number); (Thermal energy transferred)/(Fluid thermal capacity)
St_m; mass transport Station number	k_m/v	—	(Sherwood number)/(Reynolds number) (Schmidt number)
Sk; Stokes number	$l^2\Delta p/\mu\nu$	—	(Pressure force)/(Viscous force)
Sr; Strouhal number	fl/v	f = oscillation frequency, $1/T$	Used in study of unsteady flow
Su; Suratnam number	$\rho l\gamma/\mu^2$	—	(Inertia force)(surface tension force)/(Viscous force)^2
Ta; Taylor number	$\omega^2\gamma\bar{r}(r_o \quad r_i)^3/\nu^2$	r_o = outer radius, L r_i = inner radius, L $\bar{r}$ = mean radius, L	Criterion for Taylor vortex stability in rotating concentric cylinder systems
We; Weber number	$v(\rho_f l/\gamma)^{1/2}$	—	(Inertia force)$^{1/2}$ /(Surface tension force)$^{1/2}$

*In many cases, roots of, or power functions of, the groups listed here may be designated by the same names in the technical literature.

Names of quantities are followed by their dimensions. Fundamental dimensions are taken to be length [L], mass [M], time [T], temperature [θ], and electrical charge [Q]. Notation for quantities that appear only once is given in the table. The following quantities appear more than once:

C_p = specific heat at constant pressure, $L^2/T^2\theta$
d_p = particle diameter, L
D = molecular diffusivity, L^2/T
g = acceleration due to gravity = weight/mass ratio, L/T^2
h = heat transfer coefficient, $M/T^3\theta$
k_f = thermal conductivity of fluid, $ML/T^3\theta$
k_m = mass transfer coefficient, L/T
l = characteristic length, L
Δp = pressure drop, M/LT^2
t = time, T
v = charcteristic velocity, L/T
α = thermal diffusivity = $k_f/\rho_f C_p$, L^2/T
γ = surface tension, M/T^2
μ = dynamic or absolute viscosity, M/LT
ν = kinematic viscosity = μ/ρ_f, L^2/T
ρ = bubble or droplet density, M/L^3
ρ_f = fluid density, M/L^3
ρ_s = solid density, M/L^3
ω = angular velocity of fluid, $1/T$

dimensionless groups could be formed from these same variables, such as $v^3\rho/\mu g$, but these are not independent groups because, for instance, $v^3\rho/\mu g = \mathrm{Re} \cdot \mathrm{Fr}^2$. Many of these compound groups have acquired names, however, and a separate existence of their own, since any of the combined groups can be used as alternative for any one of the component groups.

F. V. A. Engel proposed that the groups forming a complete set (as defined above) for a given problem be termed true groups for that problem, and that other groups involving only independent variables in the particular problem which may be formed as combinations of the true groups be termed compound groups. He gave rules for obtaining complete sets of groups for heat transfer and flow problems, and the arguments can be readily extended to other types of problems.

Simplexes and complexes. Sometimes more than one variable of the same physical significance is important in a problem. For instance, in fluid flow through a straight rough pipe, three length parameters (pipe diameter D, pipe length L, and surface roughness ϵ) are required to define the geometry of the system. In such a case, one of these quantities (usually D in the pipe-flow problem) is used as the characteristic quantity in forming the main dimensionless groups (the Reynolds number for pipe flow), and the others occur in the form of ratios to the first (L/D, ϵ/D). These straightforward dimensionless ratios of multiple parameters arising naturally in a problem are often referred to as simplexes, while all other groups are termed complexes. Many simplexes are unnamed, being referred to simply as "geometric ratios" or the like, but a number of named simplexes exist, for example, the Knudsen and Mach numbers. Simplexes and complexes have also been termed parametric and similarity criteria, respectively.

Independent and dependent groups. Groups formed entirely from parameters which are independent variables in a particular problem have been termed independent or characteristic groups, while those which also involve dependent variables for the given problem are known as dependent or unknown groups.

Obviously, some groups which are independent in one problem may become dependent in another, according to which parameters are specified and which are unknown in each case.

In many practical cases, experimental data are correlated to give a relationship between a dependent group and one or more independent groups. Although there are certain philosophical problems involved, this procedure has led to many very useful engineering correlations. Examples are the Moody friction factor chart, which correlates the friction factor f for pipe flow in terms of the Reynolds number and the relative pipe roughness given by Eq. (1)

$$f = f(\mathrm{Re}, \epsilon/D) \tag{1}$$

and the familiar correlations for heat and mass transfer given by Eqs. (2) and (3). This type of power

$$\mathrm{Nu} = A \cdot \mathrm{Re}^{a} \cdot \mathrm{Pr}^{b} \tag{2}$$

$$\mathrm{Sh} = A' \cdot \mathrm{Re}^{a'} \cdot \mathrm{Sc}^{b'} \tag{3}$$

relationship is suggested by the Rayleigh indicial procedure for dimensional analysis. The constants A, a, and b are determined by fitting straight lines to suitable logarithmic plots of experimental data.

A special subclass of independent dimensionless groups contains only the parameters expressing the physical properties of the system being studied. This category includes such groups as the Prandtl, Schmidt, and Lewis numbers. *See* BALLISTICS; CAVITATION; CONVECTION (HEAT); DIFFUSION; FROUDE NUMBER; GAS DYNAMICS; KNUDSEN NUMBER; MACH NUMBER; PIPE FLOW; REYNOLDS NUMBER.

John Catchpole; George D. Fulford

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Dimensions (mechanics)

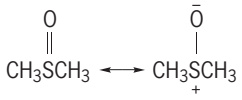
Length, mass, time, or combinations of these quantities serving as an indication of the nature of a physical quantity. Quantities with the same dimensions can be expressed in the same units. For example, although speed can be expressed in various units such as miles/hour, feet/second, and meters/second, all these speed units involve the ratio of a length unit to a time unit; hence, the dimensions of speed are the ratio of length L to time T, usually stated as LT^{-1}. The dimensions of all mechanical quantities can be expressed in terms of L, T, and mass M. The validity of algebraic equations involving physical quantities can be tested by a process called dimensional analysis; the terms on the two sides of any valid equation must have the same dimensions. *See* DIMENSIONAL ANALYSIS; UNITS OF MEASUREMENT.

Dudley Williams

Dimethyl sulfoxide

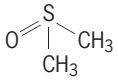
A versatile solvent, abbreviated DMSO, used industrially and in chemical laboratories as a medium for carrying out chemical reactions. Its uses have been extended to that of a chemical reagent with DMSO as a reactant undergoing a chemical change. DMSO has also gained popularity as a medicinal compound; however, its efficacy as a curative agent is not well established.

DMSO is the simplest member of a class of organic compounds which are typified by the polar sulfur-oxygen bond represented in the resonance hybrid shown in reaction (1). The molecule is pyramidal in

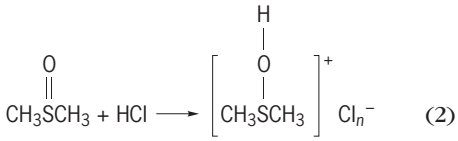


(1)

shape with the oxygen and the carbons at the corners, as in the structure below. *See* ORGANOSULFUR COMPOUND.



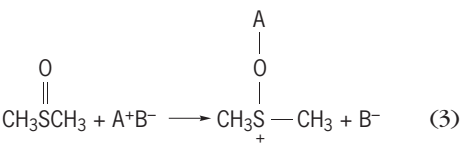
Because of the polarity of the sulfur-oxygen bond, DMSO has a high dielectric constant ($\epsilon^{20} = 48.9$) and is slightly basic. It forms crystalline salts with strong protic acids, reaction (2), and it coordi-



(2)

nates with Lewis acids. It is a colorless, odorless (when pure), and very hygroscopic stable liquid (bp 189°C or 372°F, mp 18.5°C or 65.3°F). It is manufactured commercially by reacting the black liquor, from digestion in the kraft pulp process, with molten sulfur to form dimethyl sulfide, CH_3SCH_3. The dimethyl sulfide is oxidized with nitrogen tetroxide to yield DMSO. This highly polar aprotic solvent is water- and alcohol-miscible and will dissolve most polar organic compounds and many inorganic salts.

Solvent. As an aprotic solvent, DMSO strongly solvates cations, leaving a highly reactive anion, reaction (3). Thus in DMSO, basicity and nucleophilicity



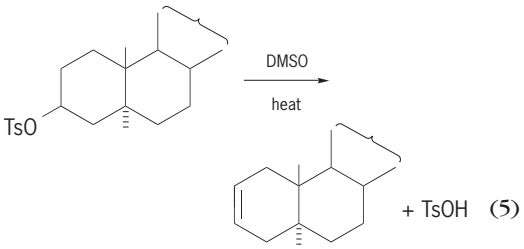
(3)

are enhanced, and it is a superior solvent for many

elimination, nucleophilic substitution, and solvolysis reactions in which nucleophile and base strength are important. Nucleophilic substitution reactions in which halogens or sulfonate esters are displaced by anions such as cyanide, alkoxide, thiocyanate, azide, and others are accelerated 1000 to 10,000 times in DMSO over the reaction in aqueous alcohol. Reaction (4) illustrates nucleophilic substitution with the

$$(CH_3)_3CCH_2Cl + NaCN \xrightarrow{DMSO} (CH_3)_3CCH_2CN + NaCl \quad (4)$$

cyanide anion. Dimethyl sulfoxide is also superior to protic solvents as a medium for elimination reactions. Reaction (5) illustrates the use of DMSO



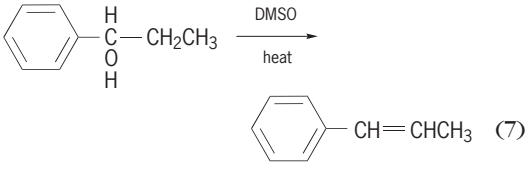
(5)

in the elimination of a tosyl group from a steroid derivative.

Eliminations can be accomplished by heating secondary halides and sulfonate esters in 100% DMSO. Other eliminations, often unsuccessful in protic solvents, are accomplished by the use of potassium *t*-butoxide [$K^{+-}OC(CH_3)_3$] in DMSO.

The elimination of tertiary and secondary alcohols has been accomplished by refluxing in 100% DMSO. Reactions (6) and (7) illustrate this elimination from a tertiary and secondary alcohol respectively.

(6)



(7)

The distinction between solvent and reactant in this and previous eliminations has not been clarified. Dimethyl sulfoxide has found use as a suitable solvent for thermal decarboxylations and base-catalyzed double-bond isomerizations. A double-bond isomerization utilizing DMSO and potassium *t*-butoxide is illustrated by reaction (8). The high dielectric constant

$$CH_2{=}CHCH_2OR \xrightarrow[DMSO]{K^{+-}OC(CH_3)_3} CH_3CH{=}CH{-}OR \quad (8)$$

makes DMSO a useful solvent for dissolving resins, polymers, and carbohydrates. It is employed commercially as a spinning solvent in the manufacture of synthetic fibers. *See* SOLVENT.

Chemical reactions. The compound is capable of taking part in several types of chemical reactions. These reactions are described below.

Coordination compounds. Dimethyl sulfoxide acts as an electron donor to form coordination compounds with a large number of inorganic ions. In most complexes the DMSO ligand is attached through the oxygen atoms, and the bond strength is roughly that of the corresponding aquo (H_2O) derivative. In addition to the typical metal-ion complexes, DMSO also forms some complexes with nonmetallic compounds. For example, BF_3 forms a one-to-one complex which is highly hygroscopic and low melting. *See* COORDINATION CHEMISTRY; COORDINATION COMPLEXES.

Synthetic intermediate. Acting as an acid, DMSO reacts as indicated by reaction (9) with sodium hydride to form the strong base methylsulfinyl carbanion. Methylsulfinyl carbanion, an effective base catalyst in many condensation reactions, itself condenses with carbonyl compounds, particularly esters, reaction (10), to give β-keto sulfoxides in high yield. The

$$(CH_3)_2S{=}O + NaH \longrightarrow \left[CH_3{-}S(=O){-}CH_2\right]^- Na^+ + H_2 \quad (9)$$

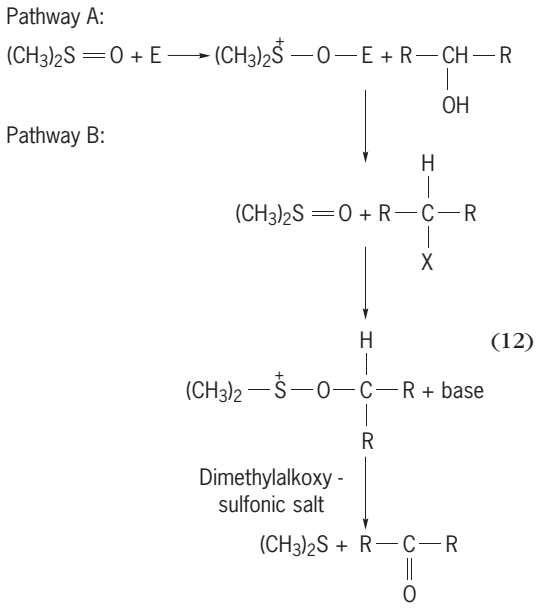
$$R{-}COOR' + CH_3SCH_2^- \longrightarrow R{-}C(=O){-}CH_2S(=O){-}CH_3 + R'O^- \quad (10)$$

synthetic importance of the easily prepared β-keto-sulfoxides lies in the fact that they can be easily reduced by aluminum amalgam to methyl ketones in a large variety of cases, reaction (11).

$$R{-}C(=O){-}CH_2{-}S(=O){-}CH_3 \xrightarrow{Al\text{-}Hg} R{-}C(=O){-}CH_3 \quad (11)$$

See CARBONYL; ORGANIC SYNTHESIS; REACTIVE INTERMEDIATES.

Oxidizing agent. Under the proper experimental conditions, many compounds (primary or secondary alcohols, halides, or sulfonates) can be oxidized to the corresponding carbonyl compound by DMSO. Most of the DMSO oxidations appear to involve the same dimethylalkoxysulfonium salt intermediate, which subsequently reacts with a base to give the observed carbonyl product and dimethyl sulfide. Through pathways A and B in reaction (12)



a compound may be converted into the dimethylalkoxysulfonium salt intermediate. The reaction pathway is determined by the structure of the substrate.

Pathway A involves reaction of DMSO with an intermediate activating electrophilic species E, which is subsequently displaced by the compound to be oxidized to form the dimethylalkoxysulfonium salt.

Pathway B involves a leaving group X (X = Cl^-, Br^-, I^-, or sulfonate) being displaced by DMSO acting as a nucleophile, and resulting directly in the dimethylalkoxysulfonium salt.

The most common technique involves use of dicyclohexylcarbodiimide as the E group and pyridine as the base and is applicable to primary or secondary alcohol groups in an almost unlimited variety of compounds including alkaloids, steroids, carbohydrates, and other complex substances. The simple experimental conditions with which most oxidations can be done have made this an important oxidation method. *See* ORGANIC REACTION MECHANISM.

Pharmacology and toxicity. The apparent low toxicity and high skin permeability have led to extensive studies in many biological systems. DMSO facilitates the transport of inorganic salts or small-molecular-weight organic compounds across skin membrane, indicating a potential hazard in commercial use.

DMSO is used as a topical anti-inflammatory agent; however, its efficacy for this purpose is not well established. Its use as such is not recommended without medical supervision because of the limited information about adverse side effects which may accompany repeated or long-term application. In some cases skin contact causes irritation and can result in lesions.

DMSO has become a readily available chemical, and many claims for its curative properties are poorly or totally unsubstantiated. Some studies failed to show any benefit of topically applied DMSO in the promotion of wound (dermal and epidermal) healing. DMSO is widely used alone or in combination with other agents to induce cell differentiation under culture conditions. Induced cellular differentiations appear to be related to DMSO-mediated alterations in cell-surface characteristics and enzyme activities. These findings suggest that DMSO should not be taken internally, taken by injection, or used topically without medical supervision.

DMSO has proven useful as a cryoprotective agent for a large number of mammalian and nonmammalian cells, probably because of its influence on membrane permeability and ability to decrease ice nucleation temperatures. Numerous physiological effects of DMSO have been studied, but its extended and routine use should be discouraged until definite clinical efficacy is demonstrated. *See* CRYOBIOLOGY.

W. W. Epstein; F. W. Sweat

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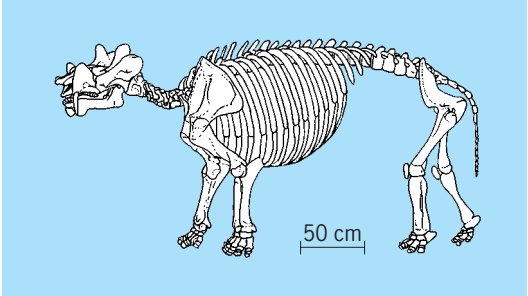
Dinocerata

An extinct order of large herbivorous mammals, often called uintatheres, from early Cenozoic deposits of North America and northern Asia. Members of this group have semigraviportal limbs, that is, adapted to bearing considerable weights, with hoofed, five-toed feet. The dentition is somewhat reduced in all forms, later ones losing all the upper incisors and in one case (*Gobiatherium*, from the late Eocene of Mongolia) even the upper canine. The cusps of the upper molars and premolars form V-shaped patterns. The lower molars and premolars possess V-shaped crests, followed by a low shelf. A saberlike canine tooth and protective lower jaw flange are present in all forms except the aberrant *Gobiatherium*, which must have relied on other means for defense. Horns are absent or very small on the most primitive forms, but by early Eocene time North American genera had begun to develop them. Middle and late Eocene uintatheres in North America developed an imposing array of six horns. One pair was on the tips of the nasal bones, another was above the root of the saberlike canine tooth, and a third pair above the ear region. *See* DENTITION; TOOTH.

The order Dinocerata consists of one family, the Uintatheriidae, which is divided into three subfamilies: Prodinoceratinae, Uintatheriinae, and Gobiatheriinae. The Prodinoceratinae includes two Mongolian genera from the late Paleocene and early Eocene, an American genus from deposits of the same ages, and a second American genus from the late Paleocene. The Mongolian genus *Mongolotherium* is the best known of these and demonstrates that the most primitive uintatheres possessed a carnivorelike body of moderate size.



Skeleton of *Uintatherium*, a middle Eocene member of the Dinocerata.

The subfamily Uintatheriinae, or horned forms, includes all other American members of the order (see **illus.**). Uintatheres increased to rhinoceros size in the Eocene, were especially common in the middle Eocene, and died out before the Oligocene began.

The Gobiatheriinae, an aberrant Asiatic side branch of the early Prodinoceratinae, is sufficiently distinct to merit subfamily rank. This subfamily, known from one late Eocene genus, is characterized by extreme reduction of the anterior dentition, including the canine, and by lack of horns. The skull is remarkably low and flat.

The Dinocerata left no descendants. They are believed to have arisen from the arctocyonid condylarths but from a different subfamily than that which gave rise to the order Pantodonta. *See* ARCHAIC UNGULATE; MAMMALIA; PANTODONTA.

Malcolm C. McKenna

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Dinoflagellida

An order of the class Phytamastigophorea, also known as the Dinoflagellata. Many botanists consider this group to be a member of the class Dinophyceae. *Noctiluca scintillans*, one of the largest species, may measure 0.06 in. (1.5 mm), whereas the flagellate state of some species may be as small as 10 micrometers. Although primarily marine, some dinoflagellates occur in fresh water. Some possess brown chromatophores, some are variously colored, and others are colorless. Masking pigments are frequent, occurring in the cytoplasm or in chromatophores. All types of nutrition exist, and some species are parasitic. Two flagella emerge laterally from a longitudinal depression or sulcus. An encircling flagellum in a groove or girdle divides the body into epicone and hypocone; the other, extending backward, propels the organism forward. The nucleus is very large. *See* CILIA AND FLAGELLA.

Dinoflagellate species, often of bizarre form, have fixed shapes determined by thick covering plates, except in Gymnodinia which are naked or have thin pellicles. Chain-formation, ameboid, and palmella aggregates occur. *Ceratium hirundinella* (**Fig. 1**) blooms in fresh water, causing tastes and odors. Ma-

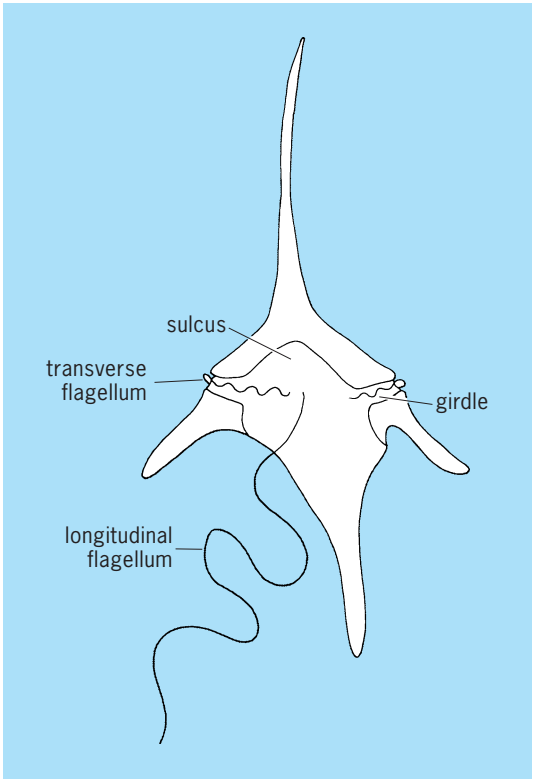


Fig. 1. *Ceratium hirundinella*, size 95–200 μm.

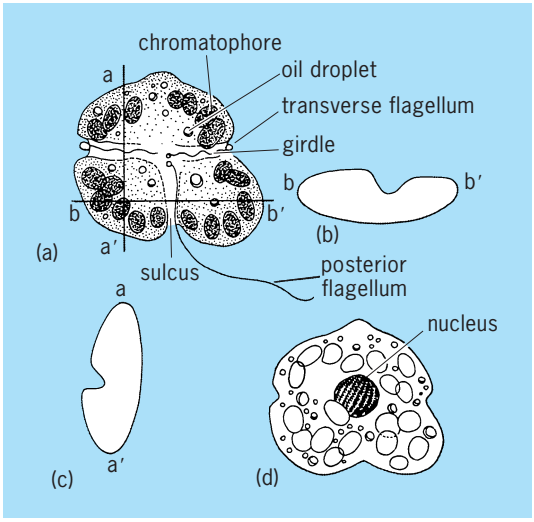


Fig. 2. *Gymnodinium breve*, which varies in size from 28 to 45 μm. (*a*) Overall view. (*b*, *c*) Transverse sections. (*d*) Fixed organism, showing central nucleus.

rine blooms of *Gymnodinium breve* (**Fig. 2**) produce the fish-killing red tides which occur along the Gulf Coast of the southern United States. *See* DINOPHYCEAE; PHYTAMASTIGOPHOREA; PROTOZOA.

James B. Lackey

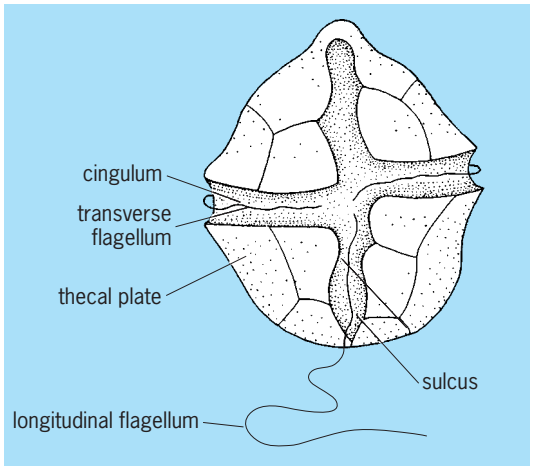
Dinophyceae

A large and extremely diverse class of biflagellate algae (dinoflagellates) in the chlorophyll *a–c* phyletic line (Chromophycota). In protozoological

classification, these organisms constitute an order, Dinoflagellida, of the class Phytomastigophora. Many taxonomists emphasize the distinctness of dinoflagellates by placing them in a separate division (Pyrrophyta or Pyrrhophyta) or even in a separate kingdom (Mesokaryota). More than 1200 species are known, classified into 18 orders and 54 families. Most are microscopic, but a few reach a diameter of 2 mm (0.08 in.). Cell shape is highly variable, with many planktonic species having elaborately modified surfaces. Dinoflagellates occur in marine, brackish, and fresh waters, frequently producing algal blooms. They may be benthic as well as planktonic, and a few are colonial. Ameboid, palmelloid, coccoid (with or without a gelatinous sheath), and filamentous forms are also known. *See* DINOFLAGELLIDA.

Structure. Most dinoflagellates at some stage in their life history bear a pair of flagella with one of two types of insertion. In one type, a posteriorly directed flagellum is located in a longitudinal groove, or sulcus, and a ribbonlike flagellum encircles the cell in a transverse groove, or cingulum (see **illus.**). The transverse flagellum, which provides the main propulsive force, comprises a striated paraflagellar rod and an undulating axoneme in a common sheath. In the other type of insertion, the flagella are borne apically or subapically on a cell that is roughly bilaterally symmetrical. One flagellum is directed forward, while the second (which has the structure of the transverse flagellum in the first type of insertion) encircles the first and beats in a perpendicular plane. Dinoflagellates frequently swim in a whirling motion, giving rise to their name (from the Greek *dinein*, to whirl).

Most dinoflagellates have an elaborate cell covering, the amphiesma, composed of several layers of concentric membranes between which may be found a cellulosic layer comprising separate plates (thecal plates). The number and arrangement of these plates have been accorded taxonomic significance. In some species the outer surface is covered by small scales.



***Gonyaulax*, a dinoflagellate. The two flagella emerge from adjacent pores in the cingulum. (*After H. J. Fuller and O. Tippo, College Botany, rev. ed., Holt, 1954*)**

Ecology. Dinoflagellates span the spectrum of nutritional diversity. About half of the species are photosynthetic, and some of these are facultatively osmotrophic (absorbing nutrients) or phagotrophic (engulfing food). Of the colorless species, some are osmotrophs while others are predatory phagotrophs. Chloroplasts are characteristically yellowish- or greenish-brown. Like those of other algae of the chlorophyll *a–c* line, they are bounded by three membranes, but differ in that the outer membrane is not part of the endoplasmic reticulum and lacks ribosomes. Chloroplast deoxyribonucleic acid (DNA) is distributed in discrete granules rather than being organized in a peripheral ring. Although photosynthetic lamellae are composed of three thylakoids, a girdle lamella is absent. The main accessory pigment is peridinin, a carotenoid unique to dinoflagellates. In some species of a few genera (*Gymnodinium*, *Peridinium*), chloroplasts are derived by an unknown process from another alga (chrysophyte or cryptophyte) and thus have uncharacteristic features. In photosynthetic dinoflagellates the main food reserve is starch, which accumulates outside the chloroplast, but storage of lipids is also common. *See* CAROTENOID; CHLOROPHYLL; PHOTOSYNTHESIS.

Members of several unrelated groups of marine dinoflagellates are parasites. The hosts are varied, comprising algae (including other dinoflagellates), protozoans, coelenterates, annelids, crustaceans, mollusks, appendicularians, and fishes. Parasitic species are highly specialized or structurally reduced, and their dinophycean affinity is often shown only by features of their motile reproductive cells and of their nucleus. Both endoparasites and ectoparasites are known.

Symbiosis involving dinoflagellates is of common occurrence in marine environments. Occasionally, colorless dinoflagellates are host to endosymbiotic algae. More commonly, dinoflagellates of the family Zooxanthellaceae are endosymbiotic in animals, including foraminiferans, radiolarians, ciliates, turbellarians, acantharians, coelenterates, and mollusks. Symbiosis is most important in coral reefs, where up to half of the carbonate in calcified structures is derived by way of photosynthesis carried out by endosymbiotic dinoflagellates.

Cell biology. The dinoflagellate nucleus is large in relation to the volume of the cell. The DNA is surrounded by a basic protein that differs from the histones of other organisms. In nearly all dinoflagellates the chromatin remains condensed, enabling the chromosomes to be seen throughout the cell cycle. During mitosis the nuclear membrane remains intact. The spindle is external, with microtubules passing through channels or tunnels in the nucleus to attach to the chromosomes. The nucleus has been termed dinokaryotic in deference to its uniqueness, or mesokaryotic with reference to its postulated intermediacy between prokaryotic and eukaryotic nuclei.

A variety of light-sensitive organelles are found in dinoflagellates. Some eyespots are free in the

cytoplasm while others are bound by membranes. In the family Warnowiaceae, complex eyelike structures called ocelli provide for the projection of an image on a retinoid. Ejectile organelles called trichocysts occur in most dinoflagellates. They are discharged as long threads when the cell is stimulated by touch, heat, or chemicals. A few members of the Polykrikaceae and Warnowiaceae have larger ejectile organelles called nematocysts because of their superficial resemblance to the nematocysts of coelenterates. A multivesicular organelle called a pusule is found near the base of the flagella of most dinoflagellates. This structure may serve an osmoregulatory or secretory function.

Bioluminescence is well known in several genera of marine dinoflagellates (*Noctiluca*, *Pyrocystis*, *Pyrodinium*, *Gonyaulax*), but its functional significance is not understood.

Reproduction. Dinoflagellates commonly reproduce by binary fission. Sexual reproduction is isogamous or anisogamous and results in the production of an encysted resting stage. The cysts are represented in the fossil record abundantly as far back as the Triassic.

Toxicity. Several dinoflagellates contain substances that are toxic to other organisms. Blooms of some of these dinoflagellates cause "red tides" that are lethal to fishes or invertebrates. Blooms of other species result in the harmless accumulation of toxic substances in the tissues of organisms that prey on the dinoflagellates. When these predators are ingested by humans, however, poisoning results. Paralytic shellfish poisoning is caused by clams or mussels that have accumulated saxitoxin produced by certain species of *Gonyaulax* or *Protogonyaulax*. Ciguatera is a poisoning caused by eating fish that have accumulated ciguatoxin derived from any of several benthic dinoflagellates. *See* ALGAE. Paul C. Silva; Richard L. Moe

Bibliography. S. P. Parker (ed.), *Synopsis and Classification of Living Organisms*, 2 vols., 1982; F. J. R. Taylor, On dinoflagellate evolution, *BioSystems*, 13:65–108, 1980.

Dinosauria

The term, meaning awesome reptiles, that was coined by the British comparative anatomist R. Owen in 1842 to represent three partly known, impressively large fossil reptiles from the English countryside: the great carnivore *Megalosaurus*, the plant-eating *Iguanodon*, and the armored *Hylaeosaurus*. They were distinct, Owen said, not only because they were so large but because they were terrestrial (unlike mosasaurs, plesiosaurs, and ichthyosaurs); they had five vertebrae in their hips (instead of two or three like other reptiles); and their hips and hindlimbs were like those of large mammals, structured so that they had to stand upright (they could not sprawl like living reptiles). Owen's diagnosis was strong enough to be generally valid, with some modification, 150 years later. His intention in erecting Dinosauria, though, as A. J. Desmond noted, seems to have been more than just the recognition of a new group: By showing that certain extinct reptiles were more "advanced" in structure (that is, more similar to mammals and birds) than living reptiles, he was able to discredit contemporary evolutionary ideas of the transmutation of species through time into ever-more-advanced forms (progressivism).

Over the next several decades, dinosaurs were discovered in many other countries of Europe, for example, *Iguanodon* in Belgium and *Plateosaurus* in Germany, but rarely in great abundance. The first dinosaur discoveries in the United States were from New Jersey as early as the 1850s [*Trachodon*, 1856; *Hadrosaurus*, 1858; *Laelaps* (=*Dryptosaurus*), 1868]. The genus *Troodon*, based on a tooth from Montana, was also described in 1856, but at the time it was thought to be a lizard. Spectacular discoveries of dinosaurs from the western United States and Canada began in a great rush in the 1870s and 1880s and continued into the early 1900s. From 1911 to 1914, expeditions into German East Africa discovered some of the largest dinosaurs ever collected: *Brachiosaurus* and *Tornieria*, among many others. In the 1920s, expeditions into the Gobi desert of Mongolia brought back new and unusual dinosaurs such as *Protoceratops*, and the first eggs immediately recognizable as dinosaurian. For various reasons, such grand expeditions dwindled until the late 1960s, when renewed activities in the western United States and Canada, Argentina, southern Africa, India, Australia, China, Mongolia, and even Antarctica uncovered dozens of new dinosaurs. The rate of new discoveries reached an estimate of one new find every 6 weeks (intriguing new ones include *Microraptor*, *Mononykus*, *Eoraptor*, and *Brachytrachelopan*) during the mid-1990s, a pace of discovery that has continued (or possibly been exceeded) in the early twenty-first century. The finds since the 1960s may represent up to 50% of all dinosaurs currently known to have existed.

Hip structure. As dinosaurs became better known, their taxonomy and classification developed, as well as their diversity. In 1887, H. G. Seeley recognized two quite different hip structures in dinosaurs and grouped them accordingly. Saurischia, including the carnivorous Theropoda and the mainly herbivorous, long-necked Sauropodomorpha, retained the generalized reptilian hip structure in which the pubis points down and forward and the ischium points down and backward. The remaining dinosaurs have a pubis that has rotated to point down and backward, thus extending parallel to the ischium; this reminded Seeley of the configuration in birds, and so he named this group Ornithischia. However, the ornithischian pubis is only superficially similar to that of birds, which are descended from, and are thus formally grouped within, Saurischia. Seeley's discovery, in fact, only recognized the distinctiveness of Ornithischia, and he concluded that Saurischia and Ornithischia were not particularly closely related. Even within Saurischia, there were general doubts that Sauropoda and Theropoda had any close

relationship; eventually, the word "dinosaur" was used mainly informally by paleontologists.

However, this situation has been reversed. In 1974, Robert Bakker and Peter Galton argued that there were a great many unique features, including warm-bloodedness, which characterized the dinosaurs as a natural group (also including their descendants, the birds). Although this scheme was debatable in some particulars, it spurred renewed studies anchored in the new methodology of cladistic analysis of phylogenies (evolutionary relationships). A 1986 analysis listed nine uniquely derived features (synapomorphies) of the skull, shoulder, hand, hip, and hindlimb that unite Dinosauria as a natural group; this analysis has been since modified and improved, and today Dinosauria is universally accepted as a natural group, divided into the two monophyletic clades, Ornithischia and Saurischia. *See* ORNITHISCHIA; SAURISCHIA.

Earliest dinosaurs. Dinosaurs are archosaurs, a group that comprises living crocodiles, birds, and all of the living and extinct descendants of their most recent common ancestor (**Fig. 1**). The closest relatives of dinosaurs, which evolved with them in the Middle and Late Triassic (about 240–225 million years ago), include the flying pterosaurs and agile, rabbit-sized forms such as *Lagosuchus* and *Lagerpeton*. The common ancestor of all these forms was small, lightly built, bipedal, and probably an active carnivore or omnivore. Somewhat larger, with skulls ranging 15–30 cm (6–12 in.) in length, were *Eoraptor* and *Herrerasaurus* from the Late Triassic of Argentina and *Staurikosaurus* from the early Late Triassic of Brazil (**Fig. 2**). When the latter two genera were first described in the 1960s, they were thought to be primitive saurischian dinosaurs. The relationships of these animals are contentious, however, with some authors placing them within Theropoda, others positioning them at the base of Saurischia, and a further group of specialists suggesting that they were outside the group formed by Saurischia plus Ornithischia. These Late Triassic genera testify to a burst of evolutionary change at this interesting time in vertebrate history, and show that a variety of taxa very close to the origin of dinosaurs appeared during this interval. The first definite ornithischians and saurischians appear at almost the same time, though dinosaurs remained generally rare and not very diverse components of terrestrial faunas until the beginning of the Jurassic Period (about 200 million years ago). *See* JURASSIC; TRIASSIC.

An area of great interest is how the dinosaurs and their closest relatives differ from their contemporaries. Their posture and gait hold some important clues. Like *Lagosuchus*, *Lagerpeton*, and their other close relatives, the first dinosaurs stood upright on their back legs. The head of the thigh bone (femur) angled sharply inward to the hip socket, which was slightly open (not backed by a solid sheet of bone as in other reptiles). The femur moved like a bird's, in a nearly horizontal plane; the shin bone (tibia) swung back and forth in a wide arc, and the fibula (the normally straplike bone alongside the tibia) was reduced

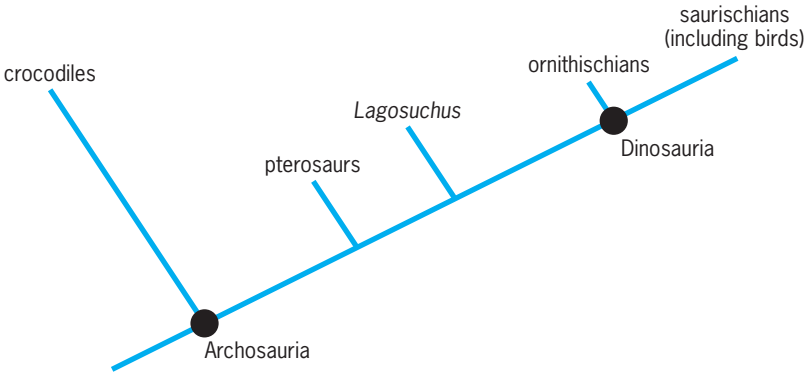


Fig. 1. Simplified phylogenetic tree of the archosaurian reptiles.

because the lower leg did not rotate about the knee, as a crocodile's or lizard's does. The ankle, too, had limited mobility: it formed a hinge joint connecting the leg to long metatarsals (sole bones), which were raised off the ground (in a stance called digitigrady, or "toe-walking"). All of these features can be seen today in birds, the living descendants of Mesozoic dinosaurs. Because the first dinosaurs were bipedal, their hands were free for grasping prey and other items, and the long fingers bore sharp, curved claws. The neck was long and S-shaped, the eyes large, and the bones lightly built and relatively thin-walled.

Ornithischia

Ornithischians (**Fig. 3**) are a well-defined group characterized by several unique evolutionary features; the entire group was analyzed cladistically in 1986, and the resulting phylogeny has been the basis of all later work. Ornithischians have a predentary bone, a toothless, beaklike addition to the front of the lower jaw that, like the front of the upper jaw, probably had a horny covering in life (**Fig. 4**). This appears to have been an adaptation for plant eating. The jaw joint was set below the occlusal plane, a nutcracker-like arrangement interpreted as increasing leverage for crushing plant material. The teeth were set in from the side of the jaw, suggesting the presence of fleshy cheeks to help sustain chewing. The cheek teeth were broad, closely set, and leaf-shaped, and were often ground down to a shearing surface. In the hip, the pubis pointed backward. In all but the most generalized ornithischians, a new prong on the pubis

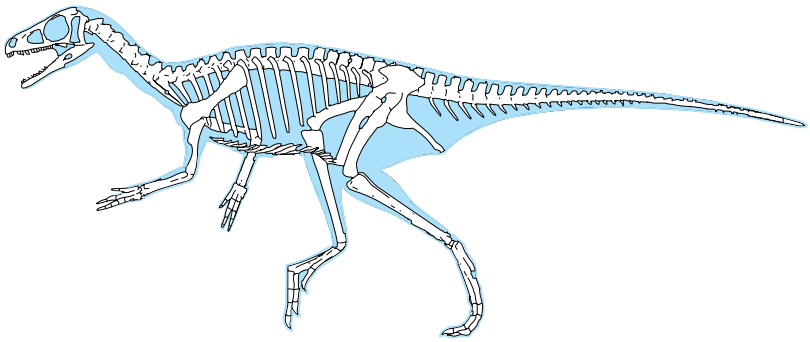


Fig. 2. *Eoraptor*, an early dinosaur or dinosaurian relative from the Late Triassic of Argentina.

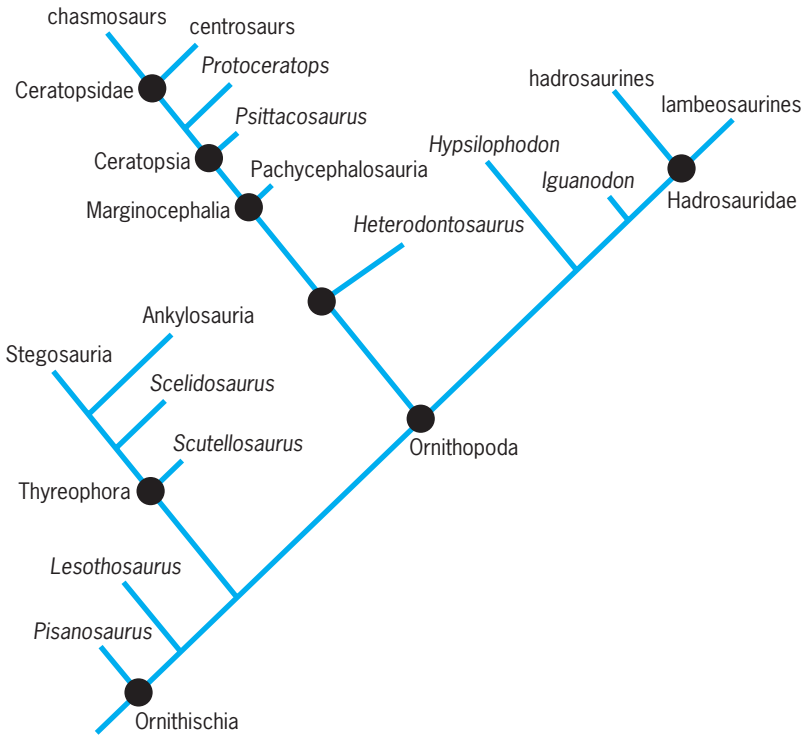


Fig. 3. Relationships of the ornithischian dinosaurs. The base of this diagram is linked to Fig. 12. (*Modified from P. C. Sereno, The evolution of dinosaurs, Science, 284:2137–2147, 1999*)

was developed from the hip socket, upward, forward, and outward. This may have provided a framework to support the guts or to anchor the hindlimb muscles.

The most generalized ornithischians known are the fragmentary *Pisanosaurus* from the Late Triassic of Argentina and the small *Lesothosaurus* from the Early Jurassic of South Africa. In the major ornithischian radiation, Thyreophora branch off first, and Cerapoda are divided into Ornithopoda and Marginocephalia. All known ornithischians are either omnivores or herbivores. *See* ORNITHISCHIA.

Thyreophora. Stegosaurs and ankylosaurs were distinguished by their elaborate armor of modified dermal bone. The earliest thyreophorans, such as the small, bipedal *Scutellosaurus* and the slightly larger, probably quadrupedal *Scelidosaurus* (both from the Early Jurassic), had such scutes all over their bodies, but not as elaborate or distinctive as those of the larger, fully quadrupedal stegosaurs and ankylosaurs.

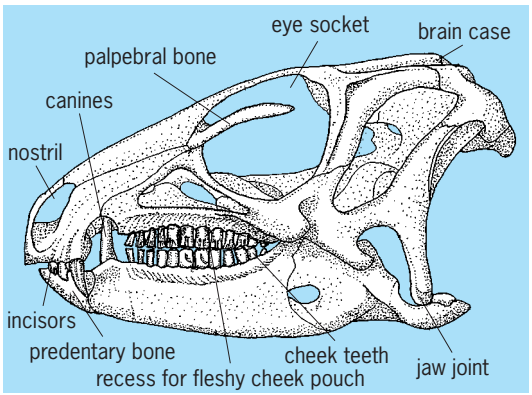


Fig. 4. Skull of *Heterodontosaurus*, showing typical ornithischian features.

Stegosaurs. Stegosaurs first appear in the Middle Jurassic of China and Europe, and were undoubtedly widespread by then, though the group never seems to have been particularly diverse: 14 genera are known, mostly from the Late Jurassic and Early Cretaceous. They had the reduced ancestral armor pattern of Thyreophora, losing the scutes on the sides of the body and retaining only a row along either side of the vertebral column (**Fig. 5**). Originally this paravertebral armor was spiky or conelike (*Huayangosaurus*), but in some forms (*Kentrosaurus*) the armor forward of the mid-dorsal region became platelike polygons, and in *Stegosaurus* the now much-enlarged, subtriangular plates occupied all but the last two terminal tail positions, which were still spiky. All well-known stegosaurs (except *Stegosaurus* itself) seem to have had a long spine projecting upward and backward from the shoulder as well. *Stegosaurus* also had a complement of bony ossicles in its throat region. In all stegosaurs the heads were relatively small and the teeth few and diamond-shaped. Stegosaurs had small brains, ranging from a walnut to a billiard ball in size. An expansion of the neural canal in the hip region was once interpreted as a "second brain" to control the hindlimbs and tail, but it is only a cavity to accommodate the sacral nerve plexus and a large glycogen ("fat") storage organ like those of birds and some other reptiles. *See* JURASSIC; CRETACEOUS.

The distinctive armor of stegosaurs has long

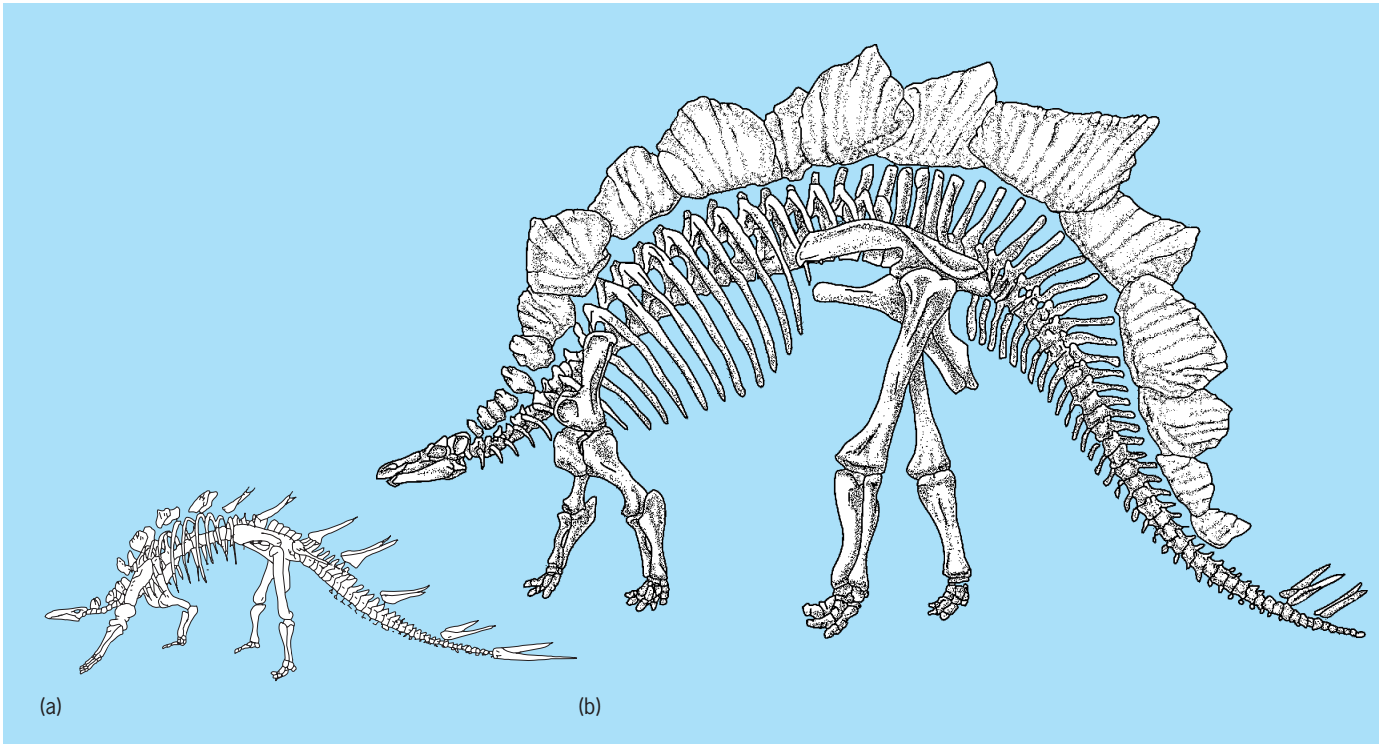


Fig. 5. Stegosaurs. (*a*) *Kentrosaurus*. (*b*) *Stegosaurus*.

fueled speculation about its position, arrangement, and function. The spikes appear to have been primitively paired, though the 17 plates of *Stegosaurus* form a single alternating, nearly medial row. Sharpey's fibers, which reflect the direction of muscle attachment to bone, show that these plates stood upright, not flat, and could not have been moved up and down. Defense is a possible function, although it must be asked why stegosaurs lost all but their paravertebral row of armor. Stegosaurs were not fast runners: the lower limbs were long and the feet and upper limbs were short, so they would have had to stand and deliver. The spikes and plates, like the shoulder spines, could have had a passive function in defense, and the terminal tail spikes, which dragged along the ground, might have made a formidable weapon. However, the plates are relatively weak "sandwiches" of latticelike bone internally, not optimal for defense. In some large stegosaurs, the broad surface area of the plates may have contributed to thermoregulation, though this hypothesis has been challenged. However, it is most likely that they were used in species recognition: no two stegosaurs have identical armor.

Ankylosaurs. Ankylosaurs are distinctive dinosaurs (**Fig. 6**). Their skulls are short, low, and flat, always

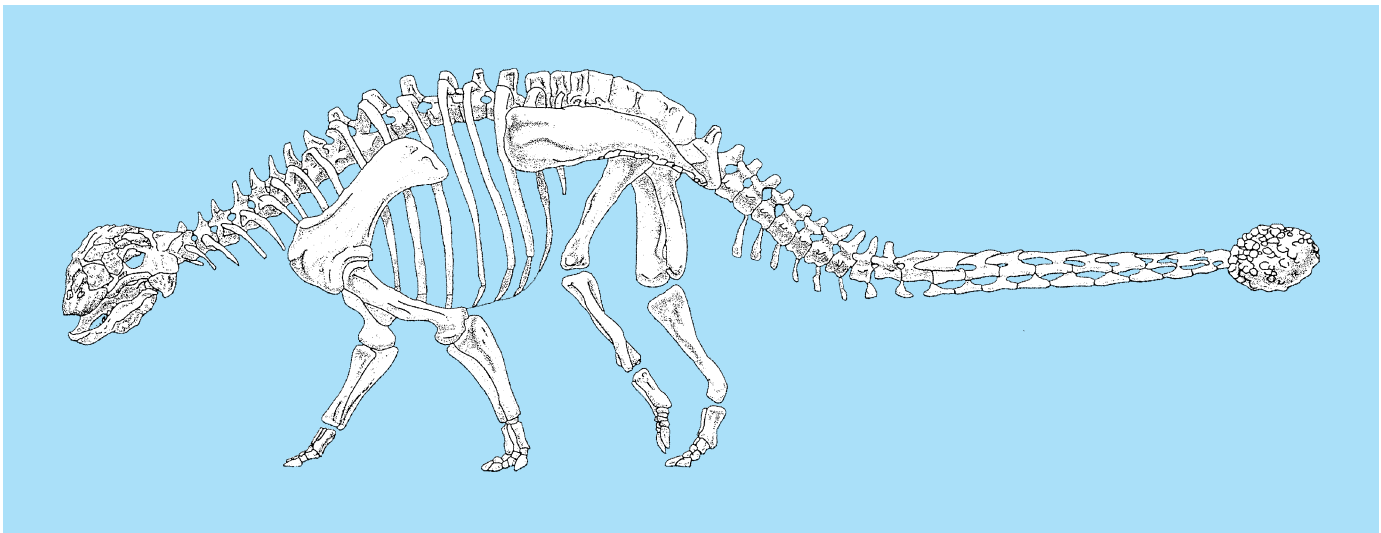


Fig. 6. *Euoplocephalus*, an ankylosaur.

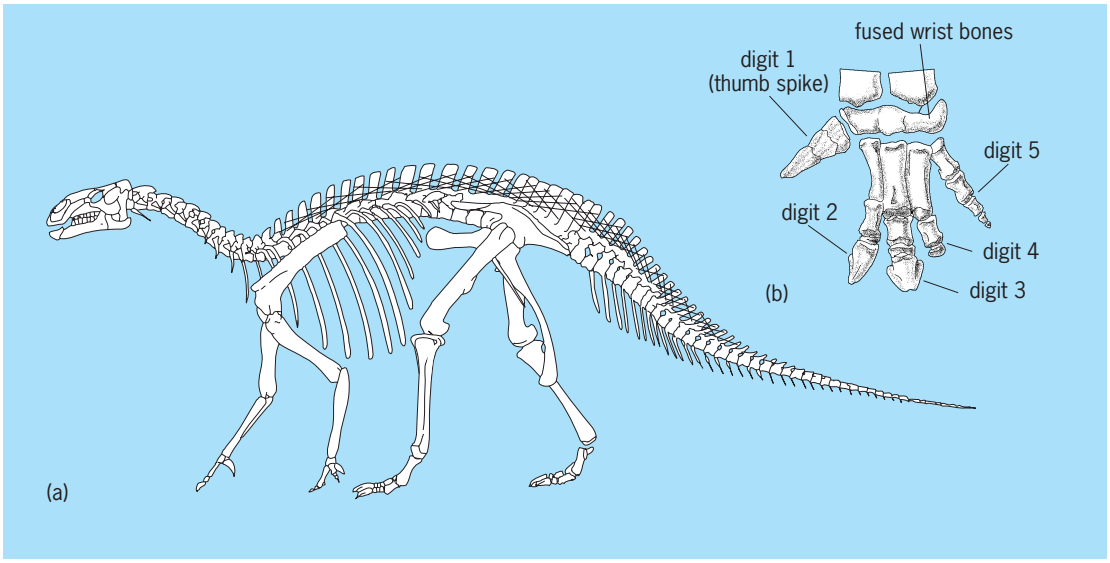


Fig. 7. ***Iguanodon*****, an ornithopod. (*a*) Skeleton. (*b*) Hand.**

wider than high in rear view, and a complement of dermal ossicles covers the skull and closes the antorbital and mandibular fenestrae. There is an S-shaped row of small, uncomplicated teeth, which sometimes display wear. The body is broad and squat, and the limbs short. Armor covers much of the body, as in Thyreophora primitively, and appears in several shapes, including keeled or spined plates, polygonal pustules, spikes, and rows of symmetrical ovals or rounded rectangles.

Ankylosauria is divided into Nodosauridae and Ankylosauridae, based on a number of distinctive but subtle features of the skull and its plating, plus several postcranial characteristics. Several authors now recognize a third group, Polacanthidae. There are not many gross differences, though ankylosaurids have a tail club (four armor plates envelop the last series of tail vertebrae, which are deeply internested and partly fused). Both groups are primarily Cretaceous, though several genera are now known from

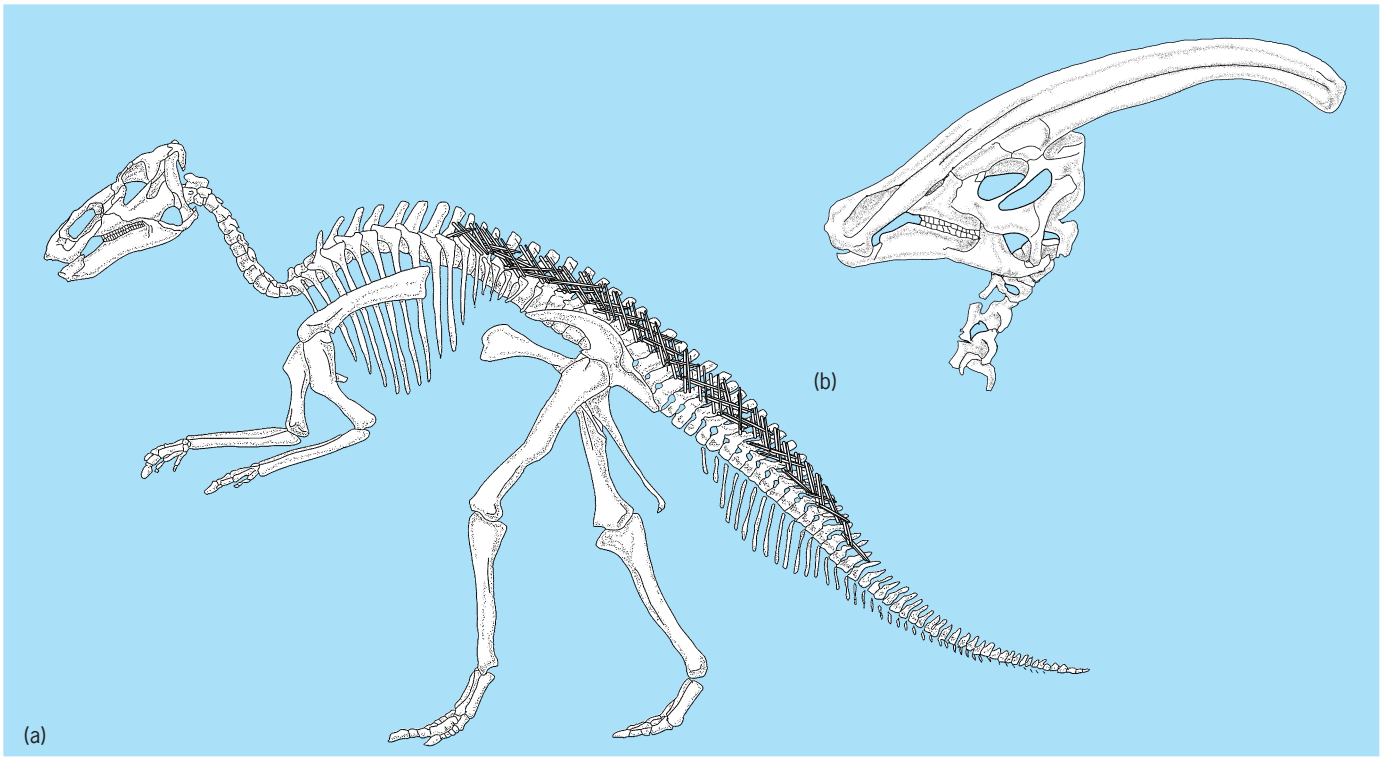


Fig. 8. Hadrosaurs. (*a*) Skeleton of the hadrosaurine *Edmontosaurus*. (*b*) Side view of skull of the lambeosaurine *Parasaurolophus*.

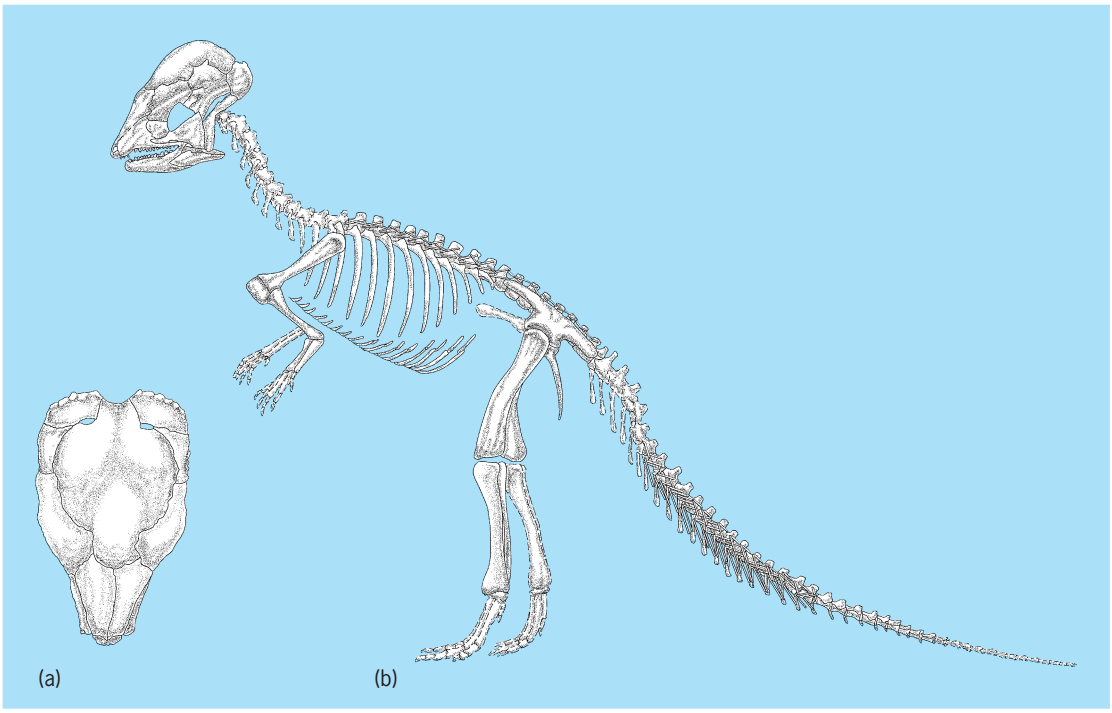


Fig. 9. ***Stegoceras*****, a pachycephalosaur. (*a*) Top view of skull. (*b*) Skeleton.**

the Jurassic. Ankylosaurs, like stegosaurs, seem never to have been very diverse, and they may have been solitary or have traveled in small family groups. They could not have been fast runners, and may have depended on camouflage, squatting, or slashing with lateral spines or tail club to repel predators.

Cerapoda: Ornithopoda. In many ways, Ornithopoda includes the most generalized ornithischians, which retain the bipedal posture of ancestral dinosaurs while lacking the horns, frills, spikes, and armor of other ornithischians. Nonetheless, they have their own specializations, and through the Mesozoic Era a general progression in size and adaptation can be seen from heterodontosaurs through hypsilophodonts to iguanodonts and hadrosaurs. In ornithopods, the upper front (premaxillary) teeth are set below the cheek (maxillary) teeth, and the jaw joint is set substantially below the occlusal plane of the teeth. *Heterodontosaurus* (Fig. 4), a fox-sized, lightly built form from the Early Jurassic of South Africa, shows these general characters and a few unusual ones, including protruding "canine" teeth presumably for defense or display.

The remaining three ornithopod groups are somewhat larger in size; they lose progressively all the front (premaxillary) teeth, reduce and/or lose the antorbital and maxillary fenestrae, and reduce the third finger (the fourth and fifth are already reduced in ornithischians). *Hypsilophodon* is a generalized member of this group. It still has five pairs of premaxillary teeth, and there is a gap between its front and cheek teeth, but the cheek teeth are now closely set and their crowns are typically worn flat. It was once thought that *Hypsilophodon* lived in trees, but this was based on the presumption that the first toe was reversed (as in birds) to serve as a perching adaptation. Actually, the anatomy of the foot is normal, and there are no additional reasons to put *Hypsilophodon* in trees.

More derived (advanced) ornithopods, such as *Dryosaurus* and *Tenontosaurus*, approach the condition of iguanodonts and hadrosaurs in many respects: the front teeth are completely lost; the front

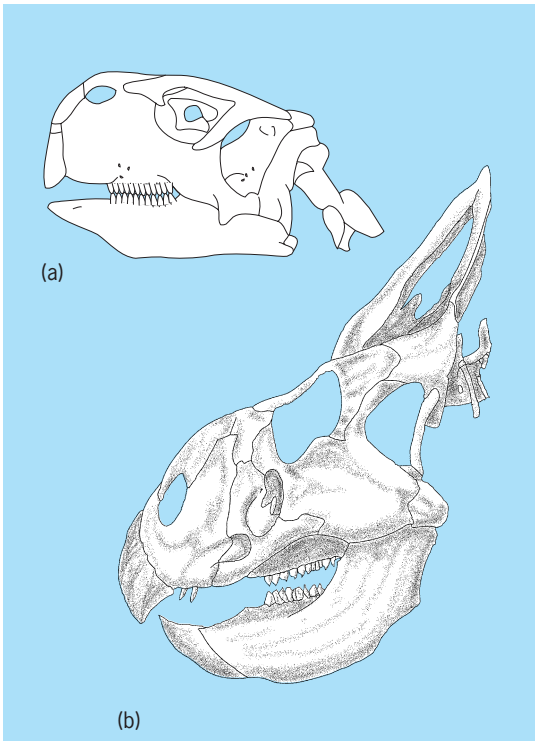


Fig. 10. Ceratopsia. (*a*) Skulls of *Psittacosaurus* and (*b*) *Protoceratops*.

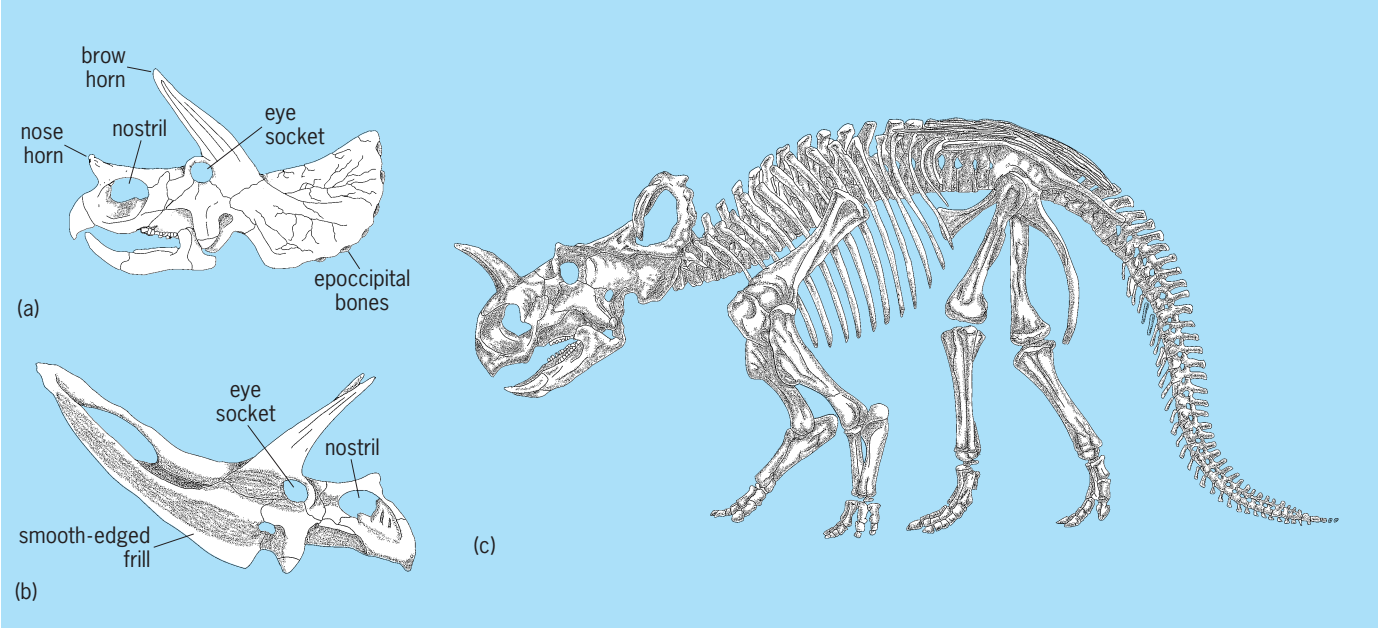


Fig. 11. Ceratopsia. (*a*) Skulls of *Triceratops* and (*b*) *Torosaurus*. (*c*) Reconstruction of *Centrosaurus*.

of the beak begins to flare; the cheek teeth become larger and more closely set, creating a single grinding battery of teeth; and the "bridge" of the nose (the area between the eyes and nostrils along the skull midline) becomes more highly arched. The forward prong of the pubis begins to grow outward until it reaches farther forward than the front of the ilium. As the cheek teeth begin to unify into opposing dental batteries, the internal kinesis (mobility) of the skull bones against each other also evolves to resist and redirect the forces of chewing.

The iguanodonts (**Fig. 7**) and hadrosaurs (**Fig. 8**) are the most familiar ornithopods, and were among the first dinosaurs discovered; they are still among the best represented in the fossil record. These two groups continue the trends already seen in the hypsilophodonts and dryosaurs, gaining larger size, losing the front teeth, flaring the snout further, and arranging the cheek teeth in a long, straight row for efficient crushing and slicing. In iguanodonts and hadrosaurs, the pubis changes somewhat: its original (backward) prong is reduced to a splint, and the forward prong enlarges and flares. This may reflect support for the intestinal tract, which is thought to have needed an extensive system of fermentation for the tough, low-nutrient plants that the animals ate. The claws on the fingers and toes are flattened and broadened into spoon-shaped hooves, and the middle three fingers, as with the middle three toes, may have functioned as a unit in walking on four legs. The fifth finger in these dinosaurs is quite distinctive: it points to the side, much like a thumb, and was similarly opposable to the other fingers, probably for manipulating food. The thumb is extraordinary in iguanodonts: the thumb phalanx, its terminal ungual, and its supporting metacarpal bone are fused together into a conelike spike, and this in turn is articulated tightly to the blocky, nearly immobile wrist bones. Whatever function this had in scratching, procuring food, or even defense, it had no function at all in hadrosaurs, which lost the thumb altogether.

Hadrosaurs (known as duckbills) dominated most of the Late Cretaceous of the Northern Hemisphere in numbers and diversity; their bones are found by the thousands in some deposits, suggesting mass deaths, perhaps during seasonal migrations or droughts. Their jaws were formidable food

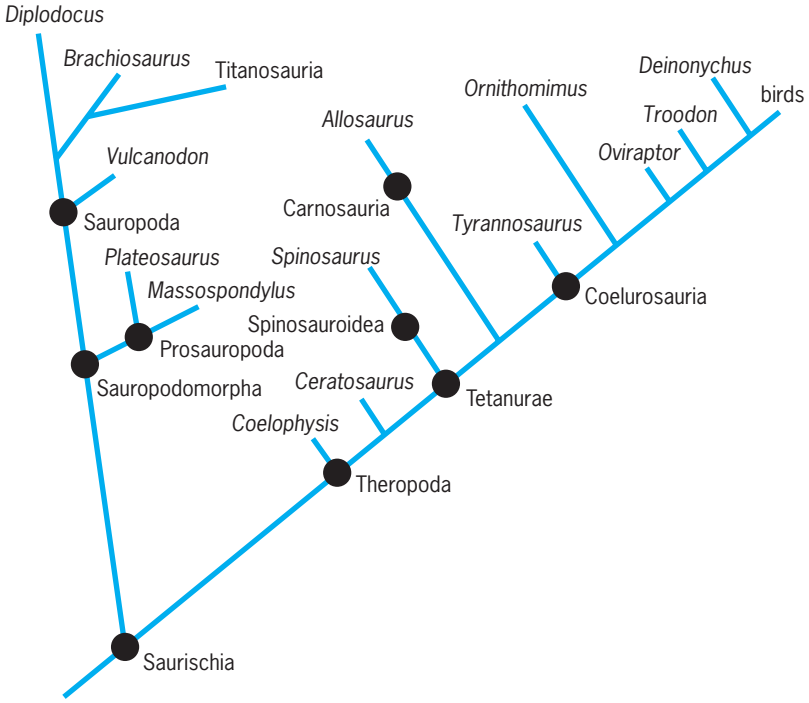


Fig. 12. Relationships of the saurischian dinosaurs. The base of this diagram is linked to Fig. 4. (*After theropod and sauropodomorph phylogenies, presented by various authors, in D. B. Weishampel et al., eds., The Dinosauria, 2d ed., University of California Press, 2004*)

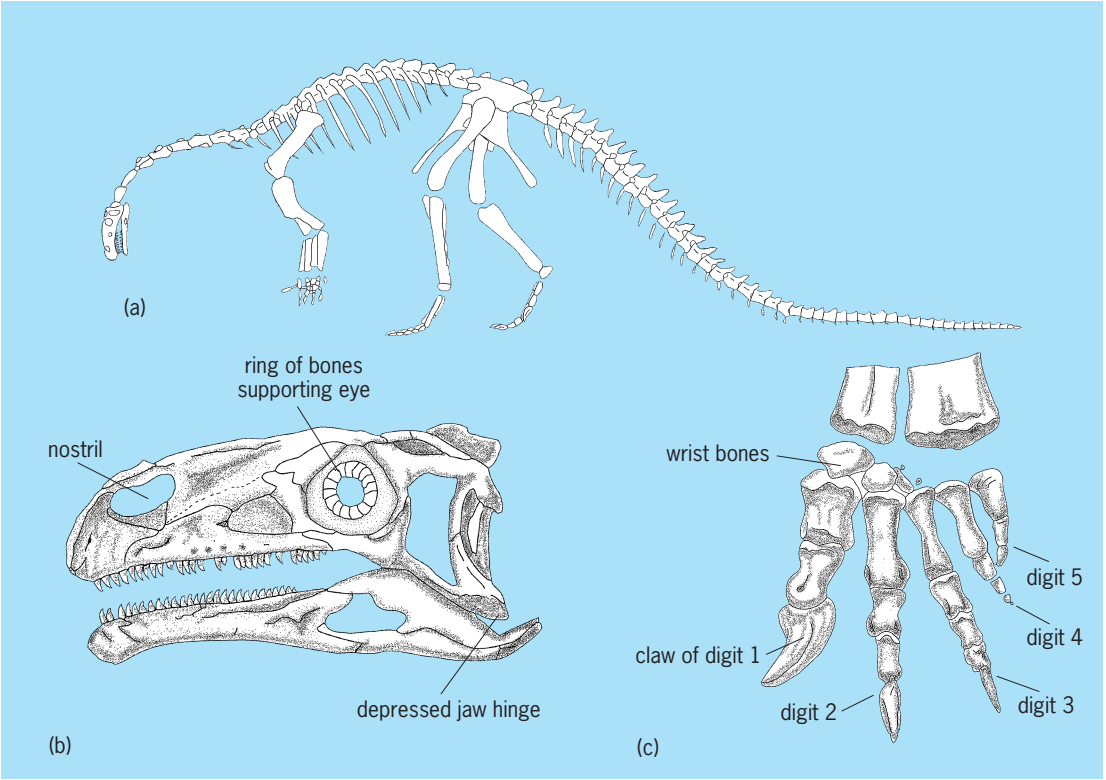


Fig. 13. Sauropodomorpha. (*a*) Skeleton of *Anchisaurus*. (*b*) Skull and (*c*) hand of *Plateosaurus*.

processors. Four or five rows of replacement teeth accompany each cheek tooth, and all of them are closely set, so as the teeth wore down they formed a single cutting and crushing surface of hundreds of teeth in each quadrant of the jaws. As in all cerapodans, the enamel was thicker on the oral side of the lower teeth and the cheek side of the upper teeth, so the teeth wore down unevenly: the cutting surface of the dental batteries was not horizontal, but beveled for better slicing as well as crushing.

Hadrosaurs are generally divided into two groups (Fig. 8): the more conservative hadrosaurines and the more derived lambeosaurines, which are noted for their profusion of skull crests and ornaments. Some crests were solid and spikelike (*Saurolophus*), while others were partly hollow and resembled bizarre hats or helmets (*Lambeosaurus*, *Corythosaurus*) or even snorkels (*Parasaurolophus*). The partial hollowness in the crests, which were formed of the premaxillary and nasal bones, was to accommodate the nasal passages that normally run through these bones. Many explanations have been proposed for this great variety in skull structure, including functions in head butting, underwater feeding, or improvement in sense of smell, but most have little support. Because lambeosaurines vary little in other skull and postcranial features, a plausible explanation is that the crests were for display and species recognition, like the horns of present-day hoofed mammals. Experimental analysis and acoustical modeling of hollow-crested forms suggested that the chambers in the crests could have served as vocal resonators to create distinctive and far-reaching sounds for species recognition and communication. This idea has contributed to the picture of sophisticated social organization that is developing for many groups of dinosaurs.

Cerapoda: Marginocephalia. Marginocephalia includes the dome-headed Pachycephalosauria (**Fig. 9**) and the frilled, usually horned Ceratopsia (**Fig. 10**). Both groups are almost exclusively Cretaceous, though a few Late Jurassic forms are now known. Marginocephalia are recognized by a small shelf or incipient frill that overhangs the back of the head; also, their hips are wider than in most other forms, though ankylosaurs have extremely wide (but very different) hips. The small shelf on the skull was the basis for the elaboration of very distinct features and functions in both Pachycephalosauria and Ceratopsia. Known from a partial postcranial skeleton from the earliest Cretaceous of Germany, *Stenopelix* is possibly an early marginocephalian.

Pachycephalosaurs (Fig. 9) are distinguished by their thick skulls, which are often ornamented with knobs and spikes. Originally, the roof of the skull was rather flat but sloped upward posteriorly and was only slightly thickened (*Homalocephale*); the two upper temporal openings on top of the skull were still prominent. In more derived forms (*Prenocephale*, *Stegoceras*, *Pachycephalosaurus*), the skull roof became rounded into a dome about the shape of a human kneecap but much thicker, and the temporal openings were covered by bone. Rows and clusters of barnacle-shaped bony knobs adorn the nose, the temporal region, and the posterior shelf in most forms, and these can be developed into long

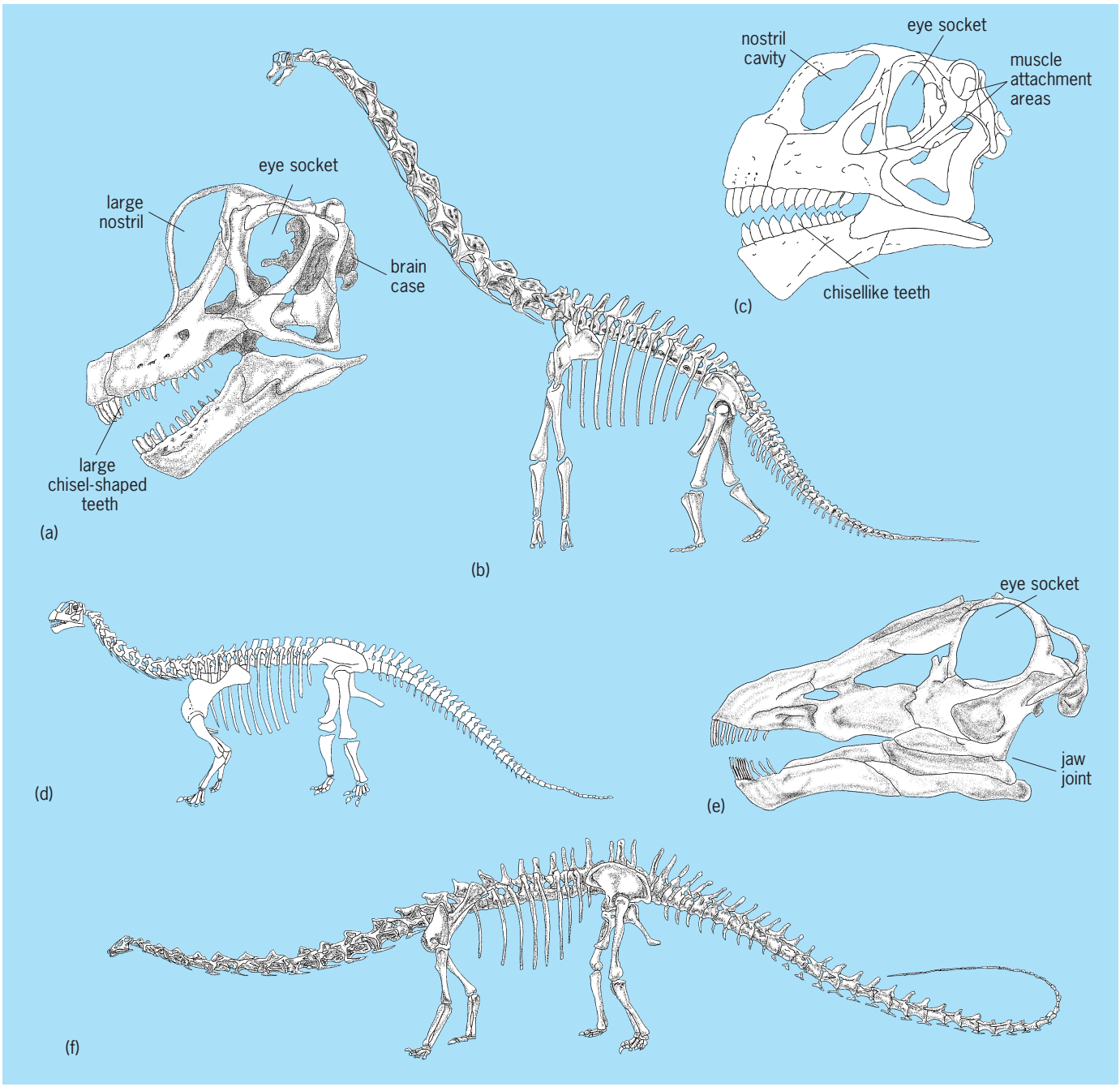


Fig. 14. Skulls and skeletal outlines of sauropods. (*a*, *b*) *Brachiosaurus*. (*c*, *d*) *Camarasaurus*. (*e*, *f*) *Diplodocus*.

spikes surrounding the dome (*Stygimoloch*). The thick, columnar bone that makes up the dome has often been interpreted as a shock absorber for head butting between competing males. However, more recent studies suggest that head-butting was quite unlikely in most forms, because the domes were so poorly designed for it. The rounded shape would cause strong torque on the neck; the dome may not have been mechanically strong enough for direct head impact; and early pachycephalosaurs had flat skull roofs that appear to have been more suited for pushing than ramming. The shape of the dome varied considerably in pachycephalosaurs, and so did its ornamentation; one pachycephalosaur had a dome with a sharp medial ridge and large spikes protruding from the sides of the head. Generally strong construction of the skull, the reinforced vertebral column (equipped with pencil-thick ossified tail tendons for support), and the strengthened pelvis suggest highly developed intraspecific (and possibly interspecific) use, perhaps mainly in merely displaying skulls to avoid direct combat.

Ceratopsia. Ceratopsia (Fig. 10), along with the hadrosaurs, were the most successful Late Cretaceous ornithischians. They are distinguished by the presence of a rostral bone, a unique, toothless, neomorphic bone that caps the front of the upper beak much as the predentary caps the lower beak in

ornithischians. Their earliest members, the basal (primitive) ceratopsians and psittacosaurs of the Late Jurassic and Early Cretaceous, were bipedal, slightly over 1 or 2 m in length (3–6 ft), and had a parrotlike beak; they were presumably browsers on low plants. The protoceratopsids initiated a trend to slightly larger size (2–3 m or 6–9 ft), quadrupedality, larger skulls, a more pronounced beak, a bony bump above the nose, and an enlarged frill at the back of the skull that usually appears as an upward-sloping arch formed by the squamosal and supporting bones.

Ceratopsidae, the great ceratopsians of the latest Cretaceous, include the centrosaurs and the chasmosaurs (**Fig. 11**). The centrosaurs generally had prominent nose horns but lacked horns above the eyes; the chasmosaurs emphasized the eye horns at the expense of the nose horns. Both groups evolved dental batteries of multiple rows of beveled teeth strikingly similar to the dentitions of hadrosaurs, and as in hadrosaurs the claws were flattened into spoon-shaped hooves. The frills were often ornamented with knobs and spikes and, as in hadrosaurs, are the most distinctive and differentiated parts of the body, the postcrania being rather conservative and not highly varied among ceratopsids. This also speaks to a function in display, rather than defense, for the great frills, which (with the notable exception of the short-frilled *Triceratops*, perhaps the best-known ceratopsian) were not solid shields but merely bony arches—hardly a suitable defense against a marauding *Tyrannosaurus* or even another ceratopsid's horns. Species recognition is a more likely use.

Saurischia

Saurischia (**Fig. 12**) includes Sauropodomorpha and Theropoda. For many years it was doubted that Saurischia formed a natural group because its two subgroups are so different, but in 1986 J. Gauthier demonstrated its validity cladistically. Uniquely derived features of Saurischia include the long neck vertebrae; the long asymmetrical hand in which the second digit is the longest; the slightly offset thumb with its short basal metacarpal, robust form, and large claw; and several other features of the skull and vertebrae. Unlike ornithischians, saurischians are fairly well represented in Late Triassic faunas as well as in the later Mesozoic. *See* SAURISCHIA; MESOZOIC.

Sauropodomorpha. This group includes the largest land animals of all time. The first sauropodomorphs (**Fig. 13**) elaborated several basic saurischian features: they had rather small skulls, elongated leaf-shaped, coarsely serrated teeth, long necks (with 10 vertebrae), robust limbs with a particularly robust thumb and claw, and a trunk that is longer than the hindlimb. *Thecodontosaurus* and *Massospondylus* typify this evolutionary stage and were probably mostly bipedal, with lengths of 2–6 m (6–18 ft). Soon, larger forms such as *Plateosaurus* appeared, and the trend toward larger size continued with the melanorosaurids of the Late Triassic and Early Jurassic. The necks became progressively longer, the bodies more massive (up to 10 m or 30 ft), and the limbs more robust, until with forms like *Vulcanodon* the condition seen in the great Sauropoda proper was attained. The relationships of these early forms, collectively called Prosauropoda, are controversial—some authors regard them as a natural group, others suggest that they are ancestral to sauropods.

Sauropoda (**Fig. 14**) includes over 100 valid genera, which are split into a number of groups. A variety of primitive forms lived during the Late Triassic to Middle Jurassic; from the Middle Jurassic onward, the more advanced neosauropods appeared. Sauropods are, in general, much larger animals than prosauropods and have at least 12 neck

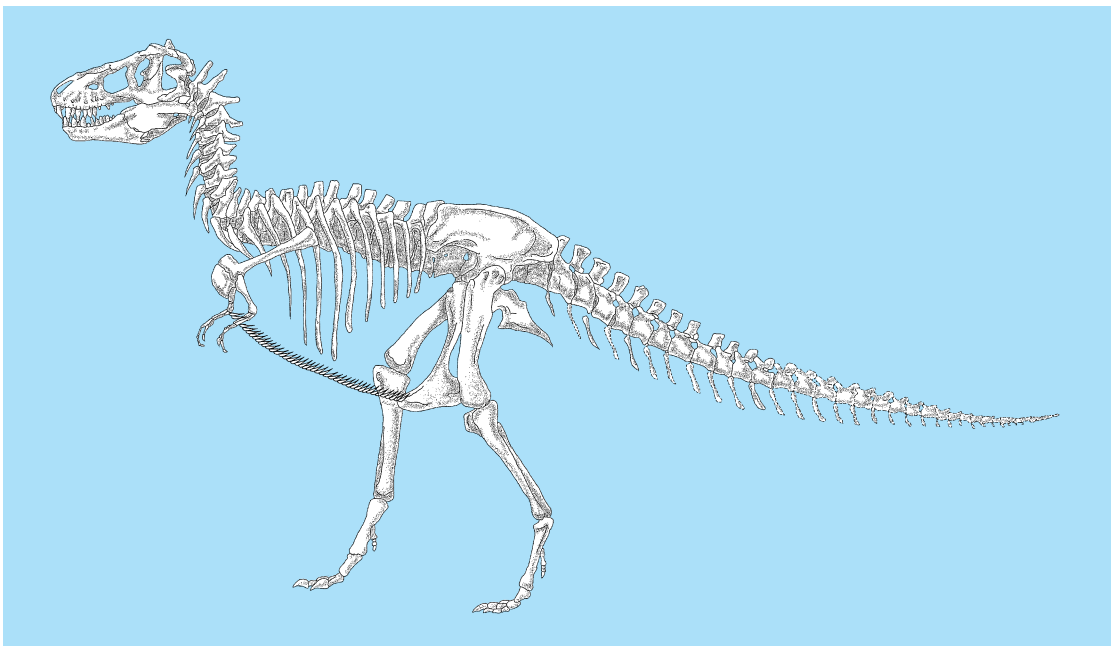


Fig. 15. ***Tyrannosaurus rex*, one of the largest known theropods.**

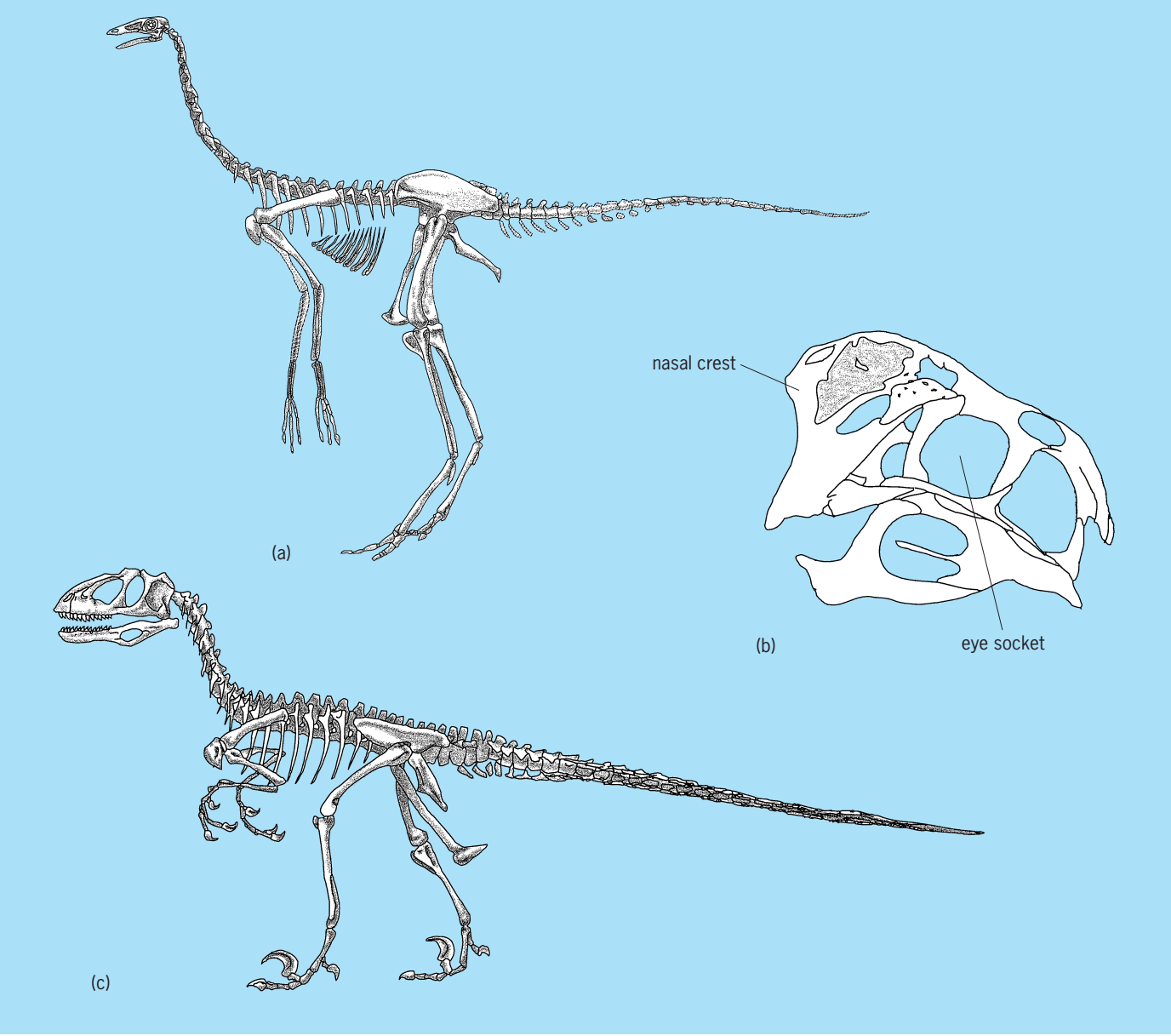


Fig. 16. Coelurosauria. (*a*) Ostrich dinosaur, *Ornithomimus*. (*b*) Skull of *Oviraptor*. (*c*) Skeleton of the dromaeosaur *Deinonychus*.

vertebrae (with a complex series of bony laminae and struts) and stout, columnar limbs. Most of the familiar sauropod groups (for example, diplodocids, brachiosaurids, and titanosaurs) are neosauropods and can be distinguished from their more primitive relatives by a number of features, including an absence of tooth serrations and a reduction in the number of wrist bones. Each sauropod group has distinctive features. For example, the diplodocids had forward-sloping teeth similar to the prongs of an iron rake, very long necks, and long, whiplike tails. About halfway along the tail, the tail chevron bones beneath the vertebrae flatten into a canoelike shape that has been interpreted as a possible adaptation for rearing up on the back legs (tripodlike) to feed in higher branches. The brachiosaurs, by contrast, had tails that were shorter than their necks, and the tail lacked the canoe-shaped chevrons of diplodocids; but their forelimbs were longer than their hindlimbs, which perhaps compensated for this lack and elevated the chest region, allowing habitual high browsing. Some sauropods, notably the titanosaur *Argentinosaurus*, could reach an estimated 70 tonnes (or 77 tons) in weight; *Seismosaurus*, a diplodocid, may have been up to 30 m (90 ft) or more in length.

Theropoda. Although meat eating seems to have been the primitive habit of dinosaurs and their close relatives, the theropods are the only group of dinosaurs to remain carnivorous. (A very small number of theropods became herbivorous.) Historically, they were divided into large carnosaurs and small coelurosaurs, virtually for convenience; but in 1986, Gauthier showed that Theropoda is a natural group united by many features of the skull and vertebrae.

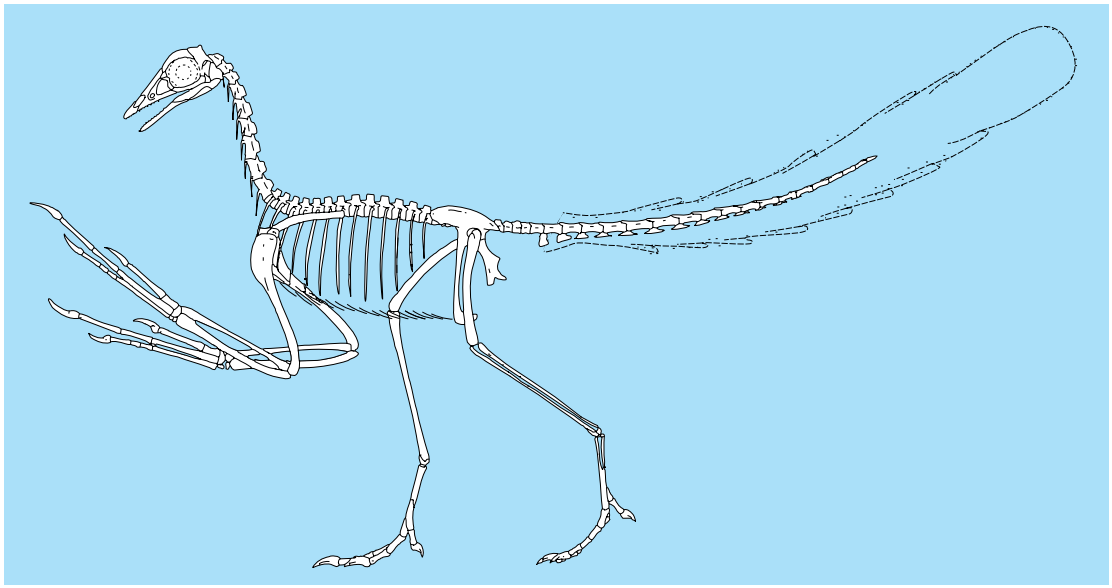


Fig. 17. Skeletal reconstruction of *Archaeopteryx*, the earliest known bird, a saurischian theropod.

In addition, the fourth and fifth fingers are lost or nearly so; the claws and teeth are sharp, and the teeth are recurved and serrated; the foot is reduced to the three middle toes; and the bones are lightly built and thin-walled.

Early in its history, Theropoda split into several lineages of primitive theropods (including coelophysoids and ceratosaurs) and a more advanced group, the Tetanurae. Primitive theropods, like the Late Triassic *Coelophysis*, were not large, but some, like the Early Jurassic *Dilophosaurus*, soon evolved to lengths of 5 m (15 ft) or more. Such ceratosaurs were the principal carnivores until the Late Jurassic. They were replaced by a great diversification of tetanuran theropods, which themselves split into three major groups: the Carnosauria (including *Allosaurus* and *Charcharodontosaurus*), the Spinosauroidea (such as *Spinosaurus* and *Baryonyx*), and the Coelurosauria (including *Ornithomimus*, *Tyrannosaurus*, *Oviraptor*, *Deinonychus*, and the birds; **Figs. 15** and **16**).

Carnosaurs had large heads with enormous, daggerlike teeth. As in nearly all theropods, the teeth were laterally compressed, with keels fore and aft for slicing, and along each keel were fine serrations that improved tearing. There is some controversy about whether large theropods were predators or scavengers. Nearly any carnivore is opportunistic enough to scavenge, and no predator takes on the strongest members of a prey species: they prefer to prey on the old, the young, and the disabled. Hence it would seem very difficult to tell the difference between adaptations for predation and those for scavenging. At any rate, few animals would want to take their chances with *Allosaurus*. Spinosauroids were similar to carnosaurs in many respects, though they possessed a number of unique specializations. For example, many forms had elongate, crocodile-like snouts, and *Spinosaurus* had tall spines on its vertebrae that might have supported a large "sail" of skin or other tissue during life.

The coelurosaurs were highly varied dinosaurs. The ornithomimids, or "ostrich dinosaurs," had reduced (or no) teeth, very long legs, and a general build reminiscent of the largest birds of today. Oviraptorosaurs had bizarrely fenestrated skulls and toothless beaks, yet the hands had sharp claws. Troodontids had long snouts with small, sharp teeth, their brains were large, and their legs and feet long. Tyrannosaurids were gigantic superpredators with massive skulls and strongly reduced forelimbs (with only two fingers per hand). The dromaeosaurs, which included the well-known *Deinonychus* and the "raptors" of the film *Jurassic Park*, had relatively large skulls and teeth, long prehensile hands and clawed fingers, a stiffened tail, and one other feature also found in some troodontids: the second (inside) toe on the foot was held off the ground and sported a trenchant, enlarged, recurved claw. In 1969 it was hypothesized that these animals would have used their long arms to grab prey, their tails to help them balance, and their large claws to rip open the bodies of their quarry. Only a few years later the Polish-Mongolian expeditions returned from the field with a pair of specimens that supported the 1969 prediction: a *Velociraptor* with its hands on the frill of a *Protoceratops*, and its feet at the prey's belly.

Dromaeosaurs were also the closest known forms to birds. The origin of birds from within dinosaurs is old: T. H. Huxley noted the strong resemblances in the 1860s. A century later J. Ostrom recognized the many unique similarities shared by *Archaeopteryx*, the earliest known bird (**Fig. 17**), and coelurosaurs, particularly forms such as *Compsognathus* and *Deinonychus*. The pubis had begun to project backward; the arms were long; the hands and fingers very long; and their digits correspond to those in the bird hand. A bony sternum had been found in some forms, and also the fused clavicles that in birds form the wishbone. The foot was

three-toed, and the first toe rotated backward as in birds. The problem was analyzed cladistically, and it was concluded that nearly 200 shared derived characters placed birds squarely within the coelurosaurian dinosaurs. In evolutionary terms, then, birds are dinosaurs, so at least one lineage of dinosaurs is not extinct. *See* ARCHAEOPTERYX.

Paleobiology

Dinosaurs laid eggs, and many nests have been found (**Fig. 18**), but matching nests and eggs to their makers is difficult unless embryos are preserved. Even then, juveniles usually lack many diagnostic characteristics of adults, so precise identification is difficult. Embryos of the hadrosaur *Maiasaura*, found in Late Cretaceous deposits from Montana, had uncalcified bones with cartilaginous joint ends, and the hatchlings were clearly helpless. Their eggshells were crushed into fragments, and the skeletons of the hatchlings found in the nests were clearly too large to have recently emerged. It appears that the young stayed around the nest and were fed by the parents until their bones fully calcified and they could fend for themselves. It is not clear how widespread either of these behaviors (or others) was among dinosaurs, but birds (living dinosaurs) are known for their extended parental care, and crocodiles, their closest living relatives, also care for their young, so it may be a general behavior among archosaurs.

Growth. If their bones are any indication, dinosaurs grew rapidly. Microscopic studies of their long bones show that the tissue was well vascularized as in birds and mammals, and that it was rapidly replaced. The texture of the bone is predominantly fibrolamellar, as in birds and mammals, and juvenile bone also shows a highly woven pattern reflecting rapid growth. Rest lines of regular occurrence have been reported in most dinosaur bones, as in those of many birds and mammals, and like these they appear to be annual. The amount of bone tissue and its many large blood vessels indicate that a lot of bone could be deposited annually, especially in juveniles and in large species. Tyrannosaurs, for example, grew twice as fast as African elephants to the same size. When dinosaurs reached full size and sexual maturity (perhaps not simultaneously), several bones co-ossified (for example, the bones that make up the braincase), as in birds and mammals. However, most dinosaur skeletons do not show these features, so they may not be fully adult.

Fig. 18. Nest of ornithischian dinosaur eggs from Montana.

Social organization. Footprints provide strong evidence that many kinds of dinosaurs traveled in groups, at least occasionally, and this is supported by records of mass burials in *Coelophysis*, *Plateosaurus*, *Allosaurus*, *Centrosaurus*, and various hadrosaurs. Social organization has the benefits of protection for the young and weak and the communication of information about food and other resources. Most major groups of vertebrates and even some invertebrates have some complex forms of social organization. Whether traveling in family groups or (at least occasionally) in large herds, all dinosaurs of any appreciable size would have had to migrate to exploit food sources successfully, if only in a local area. Annual migrations may have followed seasonal patterns of weather and vegetation.

A general picture of dinosaurian social behavior—no doubt highly variable among groups—is drawn from inferences about mass remains of trackways and bones, parental care, and the great variety of skeletal features conspicuously related to intraspecific interactions. The diversity of horns, crests, domes, knobs, and frills in many dinosaur groups contrasts starkly with the relative uniformity of their postcranial skeletons. Selection for these ornaments cannot be explained in purely functional or random terms, but strong analogies can be found in the rather similar features of living birds and mammals. Communicating, recognizing members of the same species, attracting mates, and repelling rivals, as well as delivering similar signals to members of different species, are all behavioral functions of such structures in living animals. *See* SOCIOBIOLOGY.

Metabolism. Dinosaurs evolved from within Reptilia (and so, therefore, did birds), but they are as unlike living reptiles as bats and whales are unlike horses among living mammals (and they may have been just as diverse metabolically). The physiology of extinct animals can be assessed only indirectly. No one doubts that a fossil horse was a warm-blooded, endothermic homeotherm, but such statements cannot be made about the earliest mammals and their ancestors. In the same way, the evolution of thermal strategies in dinosaurs was probably mosaic, depending on the adaptations of individual groups, and should not be considered an all-or-nothing proposition of hot-bloodedness. Many of the lines of evidence discussed above suggest that Mesozoic dinosaurs were similar in behavior and activity to mammals and birds; no evidence seems to ally them physiologically to crocodiles and lizards. But dinosaurs should be taken on their own terms, not shoehorned into models of reptile or mammal physiology based on available living analogs. The recent discoveries of nonavian dinosaurs with feathers and featherlike coverings from Early Cretaceous deposits in China indicate that the thermoregulation of at least some theropods may have been birdlike, as bone growth also shows. *See* AVES; REPTILIA.

Diversity. For 160 million years during the Mesozoic Era, dinosaurs diversified into between 500 and 1000 known genera, though many of these are based on single (often partial) specimens. Through time, dinosaurian diversity increased, but the difference between preserved diversity and some estimates of projected diversity is traced in part to varying availability of the rock record plus differential exploration. Evolution in dinosaurs was rapid: few dinosaurian genera survive more than the temporal span of a typical geologic formation (a few million years), but close relatives are often observed in succeeding formations. At the beginning of the Age of Dinosaurs, the continents were just beginning to drift apart. Dinosaurs are known from every continent and are often used to suggest land connections, such as between North America and Asia in the Late Cretaceous, or the isolation of South America during much of the Cretaceous. Some groups, such as the Ceratopsidae, apparently radiated in isolation on one or another continent.

Extinction. Many causes have been proposed for dinosaur extinction at the end of the Mesozoic Era, but most (for example, reproductive sterility, plant toxin poisoning, cataracts, supernova explosion, glaciation) have no supporting evidence. The apparent sudden decline of dinosaurs is surprising in view of their long dominance. However, the Late Cretaceous was globally unsettled in many ways. Great volcanoes were exploding in India; the Rocky Mountains were beginning to form; successive waves of regression and extinction had rocked the world's oceans; and the shallow seas were draining from many continents, including North America and Europe. These changes made the climate on land less equable and more seasonal, paving the way for destabilization of ecosystems, and perhaps creating patchy environments in which new plants and animal associations flourished. In the last few million years of the Cretaceous, known dinosaur diversity, based mainly on excellent exposures from the United States western interior, declined precipitously until, in the meters of sediment just below the Cretaceous-Tertiary boundary, only *Triceratops* and *Tyrannosaurus* survive.

An explanation is needed for this drop in diversity. Macroevolutionary patterns are based on the origination of new species and the extinction of existing ones. Dinosaur taxa were of short duration and rapid turnover; thus extinction was ubiquitous and quick throughout dinosaurian history. Toward the end of the Cretaceous, dinosaurian originations were at first burgeoning, but then they dropped; extinctions were proceeding at a near-normal pace and eventually overtook the origination rate until diversity fell to only a few taxa. In 1980, it was proposed that a giant asteroid had struck the Earth at the end of the Cretaceous, sending up a cloud of dust and vapor that blocked the sunlight for a few months or years, first lowering and then raising temperatures and causing widespread marine and terrestrial extinctions. These effects were later augmented by scenarios of global wildfires, intensely acidic rain, and other proposed environmental poisonings. As valid as the proposal of a giant asteroid impact now appears, some of the proposed biotic effects are clearly overdrawn. Dinosaurs were already declining in diversity by that time. The birds survived the impact's effects, and so did many groups of fishes, sharks, amphibians, lizards, snakes, crocodiles, and mammals. Many of these groups are notoriously sensitive to environmental disturbances. Hence any catastrophic scenarios for the terrestrial biota at the end of the Mesozoic must account both for the latest Cretaceous decline in dinosaurian diversity and the survival of any proposed environmental catastrophes by other terrestrial animals and plants. *See* EXTINCTION (BIOLOGY).

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Diode

A two-terminal electron device exhibiting a nonlinear current-voltage characteristic. Although diodes are usually classified with respect to the physical phenomena that give rise to their useful properties, in this article they are more conveniently classified according to the functions of the circuits in which they are used. This classification includes rectifier diodes, negative-resistance diodes, constant-voltage diodes, light-sensitive diodes, light-emitting diodes, capacitor diodes, and diode lasers. *See* LASER.

Rectifier diodes. A circuit element is said to rectify if voltage increments of equal magnitude but opposite sign applied to the element produce unequal current increments. An ideal rectifier diode is one that conducts fully in one direction (forward) and not at all in the opposite direction (reverse). This property is approximated in junction and thermionic diodes. Processes that make use of rectifier diodes including power rectification, detection, modulation, and switching.

In power rectification, a rectifier diode is connected in series with an alternating-voltage supply and a load. The current through the load consists of unidirectional pulses, which can be converted into essentially constant direct current by means of a filter network that removes the varying component of current. In the detection of amplitude-modulated voltage, a similar process transforms an alternating voltage of varying amplitude into a direct voltage, the magnitude of which is proportional to the amplitude

of the alternating input voltage. In amplitude modulation, diode rectification is used to vary the amplitude of a carrier voltage in response to a signal voltage. *See* AMPLITUDE-MODULATION DETECTOR; AMPLITUDE MODULATOR; ELECTRIC FILTER; ELECTRONIC POWER SUPPLY; RECTIFIER; SEMICONDUCTOR RECTIFIER.

Diode switching circuits include circuits in which output is obtained only in the presence or absence of one or more impressed control voltages. This is accomplished by the use of diodes that are normally maintained in the nonconducting (or conducting) state and are made conducting (or nonconducting) by the control voltages. Thus the diodes either provide a connection between the input and the output terminals or short-circuit the output terminals. Diodes can also be used to connect or disconnect two or more resistances in series or parallel between two terminals at selected values of input voltage and thus to synthesize a circuit that has a desired nonlinear current-voltage characteristic. *See* SWITCHING CIRCUIT.

Negative-resistance diodes. Negative-resistance diodes, which include tunnel and Gunn diodes, are used as the basis of pulse generators, bistable counting and storage circuits, and oscillators. *See* NEGATIVE-RESISTANCE CIRCUITS; OSCILLATOR; TUNNEL DIODE.

Constant-voltage diodes. Breakdown-diode current increases very rapidly with voltage above the breakdown voltage; that is, the voltage is nearly independent of the current. In series with resistance to limit the current to a nondestructive value, breakdown diodes can therefore be used as a means of obtaining a nearly constant reference voltage or of maintaining a constant potential difference between two circuit points, such as the emitter and the base of a transistor. Breakdown diodes (or reverse-biased ordinary junction diodes) can be used between two circuit points in order to limit alternating-voltage amplitude or to clip voltage peaks. Severe clipping of a sinusoidal voltage wave produces an approximately rectangular voltage wave. *See* LIMITER CIRCUIT.

Light-sensitive and light-emitting diodes. Light-sensitive diodes, which include phototubes, photovoltaic cells, photodiodes, and photoconductive cells, are used in the measurement of illumination, in the control of lights or other electrical devices by incident light, and in the conversion of radiant energy into electrical energy. Light-emitting diodes (LEDs) are used in the display of letters, numbers, and other symbols in calculators, watches, clocks, and other electronic units. *See* LIGHT-EMITTING DIODE; PHOTOELECTRIC DEVICES; PHOTOTUBE; PHOTOVOLTAIC CELL.

Capacitor diodes. Semiconductor diodes designed to have strongly voltage-dependent shunt capacitance between the terminals are called varactors. The applications of varactors include the tuning and the frequency stabilization of radio-frequency oscillators. *See* JUNCTION DIODE; MICROWAVE SOLID-STATE DEVICES; SEMICONDUCTOR DIODE; VARACTOR.

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Diopside

The monoclinic pyroxene mineral which in pure form has the formula $CaMgSi_2O_6$. Its space group is C2/c. Pure diopside melts congruently at 1391°C (2536°F) at atmospheric pressure; the effect of pressure up to 50 kilobars (5 gigapascals) on its melting temperature is given by Eq. (1), where P is in kilobars

$$P = 23.3\left[\left(\frac{T}{1665}\right)^{4.46} - 1\right] \qquad (1)$$

and T is in degrees Celsius. Diopside has no known polymorphs. Its structure consists of chains of SiO_4 tetrahedrons in which each silicon ion shares an oxygen with each of its two nearest silicon neighbors. These chains are linked together by calcium (Ca) and magnesium (Mg) ions in octahedral coordination.

Diopside forms gray to white, short, stubby, prismatic, often equidimensional, crystals with (110) cleavages intersecting at 87° (see **illus.**). Small amounts of iron impart a greenish color to the mineral. The indices of refraction increase with increasing iron; for pure diopside they are $n\alpha = 1.664$, $n\beta = 1.671$, and $n\gamma = 1.694$. Pure diopside is common and occurs as a metamorphic alteration of impure dolomites in medium and high grades of metamorphism, as shown by reaction (2).

$$CaMg(CO_3)_2 + 2SiO_2 \rightarrow CaMg(SiO_3)_2 + 2CO_2 \qquad (2)$$

Diopside shows extensive solid solution with a variety of other pyroxenes. The most important are $CaFeSi_2O_6$ (hedenbergite), $FeSiO_3$ (ferrosilite), $MgSiO_3$ (enstatite), $NaAlSi_2O_6$ (jadeite), $NaFeSi_2O_6$ (acmite), and $CaAl_2SiO_6$ (a synthetic pyroxene). Natural diopsidic pyroxenes also commonly contain significant amounts of chromium, titanium, and manganese, $CaMgSi_2O_6$ combined with variable amounts

Diopside crystals from St. Lawrence County, New York. (*Specimen from Department of Geology, Bryn Mawr College*)

of $CaFeSi_2O_6$, $MgSiO_3$, and $FeSiO_3$ form the common Ca-rich pyroxenes of basaltic and gabbroic rocks. A few natural Fe-poor diopsides containing up to 40% $MgSiO_3$ have been found in nodules in kimberlite. Natural diopsidic pyroxenes which show extensive solid solution with jadeite and to a lesser extent with acmite are called omphacite. Omphacite is a principal constituent of eclogites, rocks of basaltic composition which have formed at high pressure. *See* DOLOMITE; ECLOGITE; PYROXENE. F. R. Boyd

Diopter

A measure of the power of a lens or a prism. The diopter (also called dioptrie) is usually abbreviated D. Its dimension is a reciprocal length, and its unit is the reciprocal of 1 m (39.4 in.). Thus a thin lens of κ diopters has a focal length of $1000/\kappa$ mm or $39.4/\kappa$ in. The lens is collecting for positive κ, diverging for negative κ, and afocal for $\kappa = 0$. *See* FOCAL LENGTH; LENS (OPTICS).

One can speak of the power of a single surface. The power of a thin lens is then the sum of the powers of its surfaces in diopters. Analogously, the power of a group of (thin) lenses in contact is the sum of the powers of the single lenses.

For a lens which is not rotation-symmetric (having toric or cylindrical surfaces, for instance), two powers—one maximal and one minimal—must be assigned. These correspond to the powers in two perpendicular planes.

The dioptric power of a prism is defined as the measure of the deviation of a ray going through a prism measured at the distance of 1 m (39.4 in.). A prism that deviates a ray by 1 cm (0.394 in.) in a distance of 1 m (39.4 in.) is said to have a power of one prism diopter. *See* OPTICAL PRISM.

Spectacle lenses in general consist of thin lenses, which are either spherical, to correct the focus of the eye for near and far distances, or cylindrical or toric, to correct the astigmatism of the eye. An added prism corrects a deviation of the visual axis. The diopter thus gives a simple method for prescribing the necessary spectacles for the human eye. *See* EYEGLASSES AND CONTACT LENSES. Max Herzberger

Diorite

A phaneritic (visibly crystallized) plutonic rock having intermediate SiO_2 content (53–66%), composed mainly of plagioclase (oligoclase or andesine) and one or more ferromagnesian minerals (hornblende, biotite, or pyroxene), and having a granular texture. Diorite is the plutonic equivalent of andesite (a volcanic rock). This dark gray rock is used occasionally as a building stone and is known commercially as black granite. For a general discussion of textural, structural, and compositional differences *see* IGNEOUS ROCKS

Mineralogy. Gray or white plagioclase feldspar is the dominant mineral. The composition of the plagioclase averages in the oligoclase-andesine range (10–50% $NaAlSi_3O_8$). The plagioclase is frequently zoned, most often with a calcic core and sodic rims. Rocks with more calcic plagioclases and more abundant ferromagnesian minerals are gabbros. Subordinate light-colored minerals occurring in diorites are alkali feldspar (less than 5% of the rock). Intergrowths of alkali feldspar and quartz are common. Rocks with greater proportions of alkali feldspar are called monzonite. Those with more quartz are quartz-diorite or tonalite.

Ferromagnesian minerals in diorites consist usually of hornblende or biotite or both, but augite-, olivine-, and hypersthene-bearing diorites also occur. Biotite is more important to quartz diorites.

Accessory minerals in diorites usually include apatite, magnetite, and ilmenite, and sphene and zircon occur occasionally. Nepheline (or another feldspathoid) occurs rarely; where nephelene becomes abundant, the rock is named nepheline syenite.

Textures and structures. The texture of diorites is notably variable. Most often diorites are equigranular, with coarse, partly or mostly anhedral plagioclase and hornblende crystals, subordinate biotite, and interstitial quartz and orthoclase. Some diorites are porphyritic (diorite porphyry), with euhedral plagioclase and hornblende phenocrysts set in a granular groundmass. Some diorites contain conspicuous clots of light and dark minerals; more rarely there is a definite layering.

Occurrence. Diorite is found as isolated small bodies such as dikes, sills, and stocks, but it is also found in association with other plutonic rocks in batholithic bodies. It is closely associated with convergent plate boundaries where calc-alkalic magmatism and mountain building are taking place.

Origin. Many diorites are hybrid rocks that are formed by reaction of magma with wallrocks, or by mixing of two magmas. Most diorites crystallized from a magma which was dioritic or andesitic, perhaps differentiated from a basaltic melt. *See* MAGMA. William Ingersoll Rose, Jr.

Dioscoreales

A widespread order of lilioid monocotyledons composed of four or five families, the most important of which is Dioscoreaceae (650 species). These families contain some of the most peculiar plants among the monocots. Two families, Thismiaceae and Burmanniaceae (with 35 and 140 species, respectively), are largely composed of nonphotosynthetic herbaceous species that parasitize fungi, whereas other members, such as the bat plant (*Tacca*, Taccaceae) and the yams (*Dioscorea*) have net-veined leaves similar to dicots. Most of the 600 species of *Dioscorea* are vines, and all contain diosgenin, a precursor of progesterone and cortisone, leading to their being commercially collected in some parts of the world. *See* MAGNOLIOPHYTA. Mark W. Chase

Diphtheria

An acute infectious disease of humans caused by *Corynebacterium diphtheriae*. Classically, the disease is characterized by low-grade fever, sore throat, and a pseudomembrane covering the tonsils and pharynx. Complications such as inflammation of the heart, paralysis, and even death may occur due to exotoxins elaborated by toxigenic strains of the bacteria.

Etiologic agent. *Corynebacterium diphtheriae* is an obligate aerobic gram-positive bacillus that is shaped like a club. Whether a strain produces toxin or not determines the severity of the disease.

Toxigenicity is mediated by both genetic and nutritional factors. The gene coding for toxin production, tox^+, resides in several bacterial viruses (phages). When nontoxigenic *C. diphtheriae* are infected (lysogenized) by these phages, they are converted to toxin-producing organisms. In vitro, a low iron concentration in media is required for toxin production; the clinical significance of this nutritional requirement is unknown. *See* TOXIN.

Pathogenesis of diphtheria is mediated not only by the bacterial toxin but also by its cell-wall antigens and other enzymes. Nontoxigenic strains, therefore, can also produce localized disease, albeit a milder form.

Epidemiology. Humans are the only natural hosts of *C. diphtheriae.* Transmission is person to person, mostly by intimate respiratory and physical contact; the latter is more important in cutaneous diphtheria. Immune persons can become asymptomatic carriers of toxigenic *C. diphtheriae.* In the prevaccine era, respiratory diphtheria ranked among the leading causes of morbidity and mortality among children, presumably because of their high contact rates. Peaks in incidence occurred approximately every 10 years, with marked winter seasonality in temperate zones. The sexes were affected equally. In the tropics, cutaneous diphtheria predominates and is not seasonal. In unvaccinated populations, 75% or more of the population is likely to develop immunity to diphtheria by adulthood, presumably from mild or subclinical infections.

Where childhood vaccination against diphtheria is routine, the disease has practically disappeared. In North America, diphtheria (mostly cutaneous) is occasionally seen among Native Americans and indigent urban dwellers. The incidence has declined disproportionately to the fraction of the population with protective antitoxin titers. The absence of this disease may be due to an indirect effect of vaccination in reducing the circulation of toxigenic strains. Vaccine-induced immunity will wane in the absence of revaccination or exposure to toxigenic strains, leading to increased susceptibility among adults. When susceptibility increases and toxigenic strains are reintroduced, however, diphtheria outbreaks can recur. Adult susceptibility, reintroduction of toxigenic strains, and poor socioeconomic conditions led to a massive epidemic of diphtheria in the former Soviet Union, with more than 150,000 reported cases and over 5000 deaths from diphtheria from 1990 through 1998. Most reported patients in this outbreak have been 15 to 50 years old.

Signs and symptoms. Patients with pharyngotonsillar diphtheria experience sore throat, difficulty in swallowing, and low-grade fever. Laryngo-tracheal diphtheria results in hoarseness, croupy cough, difficulty in breathing, and pallor. Persons with severe cases may develop stupor, coma, and death from circulatory or respiratory failure within 10 days. Nervous system effects, principally motor, may appear 3–7 weeks after the onset of disease.

Diagnosis. Even if the classic diphtheritic membrane is present in the throat, clinical diagnosis is difficult, as a pseudomembrane also occurs in infectious mononucleosis, Group A streptococcal pharyngitis, and candidiasis. Specific diagnosis requires culture isolation of *C. diphtheriae* from the diseased site. Growth generally requires 8–12 h if no antibiotics are administered. Follow-up testing for toxigenicity is critical if *C. diphtheriae* is isolated.

Pathogenesis. The upper respiratory tract is the most common portal of entry for *C. diphtheriae*. It can also invade the skin and, more rarely, the genitalia, eye, or middle ear. The disease has an insidious onset after a usual incubation period of 2–5 days. Clinical severity is determined by the individual's preexisting immunity against diphtheria toxin, the virulence and toxigenicity of the invading *C. diphtheriae*, other coexisting disease, and the anatomic location of the infection. Anterior nasal and cutaneous diphtheria tends to be mild. Tonsillar diphtheria may extend either upward to involve the soft palate and other parts of the pharynx, or downward into the larynx or bronchi.

Toxin. Diphtheria toxin is a 60-kilodalton polypeptide that inhibits protein synthesis by blocking adenosine diphosphate (ADP) ribosylation and causes apoptotic cell death by a separate but related mechanism. The toxin is released by the multiplying *C. diphtheriae* and is absorbed by mucosal epithelial cells; their death elicits an inflammatory response. When the fluid that is released from the tissues clots, it enmeshes fibrin, dead epithelial cells, leukocytes, and bacterial organisms to form a classic adherent diphtheric pseudomembrane 1–3 mm thick. The membrane has a leatherlike appearance, and bleeding occurs when it is removed. It may be bluish, white, gray, or black, depending on the degree of hemorrhage. The membrane becomes a fertile site for further bacterial growth and toxin production. Its extension creates a larger surface for toxin absorption and portends a grave prognosis. Absorbed toxin is highly toxic to cardiac and nerve cells, leading to the common complications of myocarditis and polyneuritis.

Complications. Persons with severe pharyngeal diphtheria develop marked edema of the anterior neck resulting in a "bullneck" appearance. In laryngeal or bronchial diphtheria, the membrane provides

a large surface for toxin absorption and can suddenly occlude the airway. Absorbed toxin can damage any human tissue, but affects the heart and nervous system most severely. The resultant fatty degeneration of cardiac muscle can lead to arrhythmias, dilatation, and heart failure. Nerve damage, usually delayed and transient, can result in paralysis of the soft palate, posterior pharyngeal wall, cranial nerves, and respiratory muscles. Fatalities for diphtheria before the advent of antitoxin therapy occurred in 35% of cases overall, and were as high as 90% for laryngeal diphtheria; even with antitoxin therapy, the fatality rate is still approximately 5–10%.

Treatment. The only specific therapy is diphtheria antitoxin, administered in doses proportional to the severity of the disease. Antitoxin is produced by hyperimmunizing horses with diphtheria toxoid and toxin. It is effective only if administered prior to the binding of circulating toxin to target cells. Therefore, early use of antitoxin is advised in suspected cases with an inadequate or unclear history of immunization. Antibiotics do not alter the course, the incidence of complications, or the outcome of diphtheria, but are used to eliminate the organism from the patient. This helps to prevent further toxin production and transmission of the organism. Erythromycin or penicillin is preferred. Patients should be isolated until two successive cultures, taken at 24-h intervals, are negative. The patients should be carefully observed for cardiac and respiratory complications.

Prevention. Persons with protective antitoxin titers may become infected with diphtheria but do not develop severe disease. Since the 1920s, active immunization with diphtheria toxoid has proved safe and effective in preventing diphtheria in many countries. Diphtheria toxoid is produced by incubating the toxin with formalin. It is usually administered together with tetanus toxoid and pertussis vaccine as DTP. Active immunization requires a primary series of four doses, usually at 2, 4, 6, and 18 months of age, followed by a booster at school entry. Older persons should receive adult tetanus and diphtheria toxoids to complete their primary series, and boosters every 10 years. *See* IMMUNITY.

Control efforts following diagnosis of an initial case should be directed toward the early diagnosis, isolation, and treatment of cases; full immunization of children and adults; and antibiotic treatment of close contacts. *See* MEDICAL BACTERIOLOGY; VACCINATION.

Robert Chen; Charles Vitek

Diplogasterida

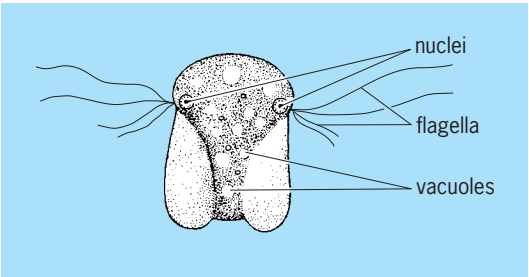
An order of nematodes in which the labia are seldom well developed; however, a hexaradiate symmetry is distinct. The external circle of labial sensilla may appear setose, but they are always short, never long or hairlike. The stoma may be slender and elongate, or spacious, or any gradation between. The stoma may be armed or unarmed; the armature may be movable teeth, fossores, or a pseudostylet. The corpus is always muscled and distinct from the postcorpus, which is divisible into an isthmus and glandular posterior bulb. The metacorpus is almost always valved. The female reproductive system may have one or two ovaries, and males may or may not have caudal alae; however, a gubernaculum is always present. The male tail commonly has nine pairs of caudal papillae; three are preanal and six are caudal.

There are two superfamilies in this order, Diplogasteroidea and Cylindrocorporoidea. The members of Diplogasteroidea are predators, bacterial feeders, and omnivores; they are often found in association with insects or the fecal matter of herbivores. The Cylindrocorporoidea includes both free-living forms and intestinal parasites of amphibians, reptiles, and certain mammals. *See* NEMATA.

Armand R. Maggenti

Diplomonadida

An order of the class Zoomastigophorea, phylum Protozoa. These are small, colorless flagellates, some of which are free-living and some parasitic. The body of this protozoan is bilaterally symmetrical, composed of two mirror halves, each with a nucleus and a full set of organelles. There are four flagella to a side, not all of the same length. In division of the cell the two nuclei each form a spindle, so that each daughter cell receives two nuclear complexes.



A diplomonad, *Trepomonas rotans*.

Trepomonas rotans is probably the most common of the free-living species (see **illus.**), and usually occurs in water of high organic, low (or no) oxygen content, such as sewage. It also occurs in ocean water. This protozoan may have two or four long lateral flagella and six or four short ones, according to the species. *Trepomonas* swims slowly with jerky movements, revolving on its axis. This genus also is found as a parasite or symbiont in other animals, such as amphibians. *Hexamita* is another genus containing both free-living and parasitic representatives. *Spironucleus*, named for its spiral nuclear arrangement, is parasitic. *Giardia* has several species which parasitize vertebrates. In humans giardiasis results in severe dysentery; however, adequate medication is available. *See* GIARDIASIS; PROTOZOA; SARCOMASTIGOPHORA; ZOOMASTIGOPHOREA.

James B. Lackey

Diplopoda

A class of terrestrial, tracheate, oviparous arthropods. Diplopods (millipeds) are largely cryptic in habits, are saprophytic feeders, and are characterized by the development of a compact head with a pair of short, simple, eight-jointed antennae, powerful mandibles, and a subbuccal gnathochilarium formed from embryonic maxillary elements. The body is not differentiated into thorax and abdomen, but is composed of a variable number of similar, cylindrical diplosomites, each of which (except the first two or three) bears two pairs of walking legs (see **illus.**). The body wall is chitinous, with a thick impregnation of calcium carbonate in the majority of species. Most millipeds are variously adapted for rolling into a closed spiral or nearly perfect sphere when threatened. Typically, each abdominal segment is provided with a pair of glands which effuse a volatile poisonous fluid. Respiration occurs through profuse fine tracheae opening through stigmata near the bases of the legs. The digestive tract is a straight tube, except in oniscomorph species in which it is somewhat coiled. The anal opening is in the last segment and is closed by two tightly fitting anal valves. Ducts from reproductive organs pass through or behind the second pair of legs to external openings.

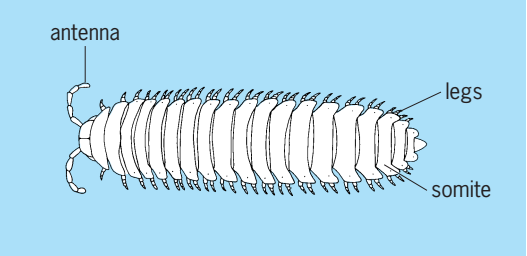
Reproduction. The sexes are separate and fertilization is internal, following prolonged clasping behavior. Spermatic masses are extruded beforehand from the male seminal opening onto the gonopods (modified legs of the seventh segment) from which sperm material is transferred into seminal receptacles in the cyphopods (specialized structures terminating the outer ends of the oviducts). In oniscomorph millipeds the gonopods are not developed, and the male transfers spermatophores with his mouthparts. Eggs vary greatly in size and number. They may be laid in a cluster and "brooded" by the mother, scattered singly in the humus environment, or enclosed in an igloo-shaped mud nest built by the mother. Postembryonic development is anamorphic, and the individual passes through seven growth stages, with segments and legs added at each molting period. Development is gradual, without major changes in appearance, and may require a year or more for completion. Mating usually takes place soon after the molt into sexual maturity.

Classification and distribution. More than 8000 species have been described, although so far only the fauna of Europe is well known. Existing classifications are still unsatisfactory but, in general, about 11 orders and more than 111 families are recognized. Probably as many as 25,000 species will eventually be recognized. Classification is based to a large extent upon shape of the male gonopods, which are quite constant and characteristic for each species. For higher groups the forms of segments, mouthparts, and legs are useful.

Most species are local in distribution, because millipeds generally remain close to the parental habitat, and some may be restricted to a few square miles. Even genera are limited in distribution, and only a few occur on more than one continent. This endemism, plus dependence on a continuously moist and undisturbed habitat, makes diplopods well suited for studies of zoogeography and evolution. *See* ZOOGEOGRAPHY.

The diplopods have a long geological history; they arose in the Early Devonian and were well developed by the Late Pennsylvanian. There has been little adaptative radiation in the Diplopoda, and less external modification of body form than in perhaps any other large class of arthropods. *See* ARTHROPODA.

Richard L. Hoffman



A diplopod. (*After R. E. Snodgrass, A Textbook of Arthropod Anatomy, Cornell University Press, 1952*)

Diplura

An order of ancestrally wingless Hexapoda comprising about 800 species and variously considered as apterygotan insects or as their independent sister group. These insects are equipped with three pairs of thoracic legs and a pair of antennae on the head. They are slender and white, but unlike typical insects they have intrinsic muscles in the antennal segments beyond the basal one. All diplurans lack eyes, and have paired mandibles and maxillae largely enclosed by fused folds of the head so that only the tips are exposed for retrieving food. The integument is thin and unpigmented, and the posterior end of the abdomen is furnished with a pair of cerci, which may be multisegmented (for example, in the family Campodeidae) or single segmented and modified into a pair of stout, pigmented jaws or forceps (in the family Japygidae and allied families) resembling those of adult earwigs and used for catching prey or for defense. The dipluran body size is usually small, under 0.4 in. (1 cm) long, but some Australian *Heterojapyx* species reach 2 in. (5 cm).

Their habitat is normally cryptic in soil, under stones, and in rotten wood, where Campodeidae feed on vegetable matter, while other families apparently are predaceous. Campodeidae are found worldwide and are common in temperate and tropical regions, while the other eight families are mainly restricted to warmer climates and are often discontinuous or local in distribution. *See* INSECTA.

William L. Brown, Jr.

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Industrial Research Organization Staff, *The Insects of Australia*, 2d ed., 1991.

Dipnoi

The infraclass, also called Dipneusti, consisting of the lungfishes. The Dipnoi plus the fossil infraclass Porolepimorpha make up the subclass Dipnotetrapodomorpha, one of five subclasses of the class Sarcopterygii. The extant lungfishes are in two orders as follows:

Order Ceratodontiformes
  Family Ceratodontidae (Australian lungfishes)
Order Lepidosireniformes
  Family Lepidosirenidae (South American lungfishes)
    Protopteridae (African lungfishes)

In comparison with Devonian lungfishes, extant species have a reduced number of median fins, different tail shape, and decreased ossification of the braincase and other endochondral bones. For example, in Devonian lungfishes the caudal fin is heterocercal (that is, the dorsal lobe is usually longer) or diphycercal (symmetrical), whereas it is always diphycercal in extant species; the dorsal and anal fins are free from the caudal fin, whereas in extant species the dorsal and anal fins are confluent with the caudal fin; there are zero to three branchiostegal rays, whereas in extant species branchiostegal rays are always absent; and a gular plate is present, whereas in extant species it is absent. Also, cell size has apparently increased since the Devonian, and the living species have very large genomes. These changes suggest that pedomorphosis played a role in dipnoan evolution. *See* HETEROCHRONY.

Recent forms. Lepidosireniforms have eel-like bodies, filamentous paired fins, two lungs, tooth plates with three cutting blades, a greatly reduced skull roof, and an elongated ceratohyal important in suction feeding and respiration (**Fig. 1**). Lepidosirenids are obligate air breathers, and the heart, arterial tree, and gills are highly specialized. To breathe air, the fish protrudes its snout from the water surface, gulps air into the mouth, and then forces it into the lung by closing the mouth and raising the floor of the oral cavity. Male lepidosirenids construct and defend breeding nests containing thousands of young. *Lepidosiren paradoxa* occurs in the Amazon and Parana river systems in South America. Young South American lungfish have yellow spots, and are imported to the United States for the tropical fish trade.

African lungfishes of the genus *Protopterus* are widely distributed in major lake and river systems of sub-Saharan Africa. Four species are known: *P. annectens*, *P. dolloi*, *P. aethiopicus*, and *P. amphibius*. Some reach nearly 6 ft (2 m) in length and are important in local fishing economies. African lungfish survive periods of drought by burrowing in mud and encasing themselves in cocoons of hardened mucus.

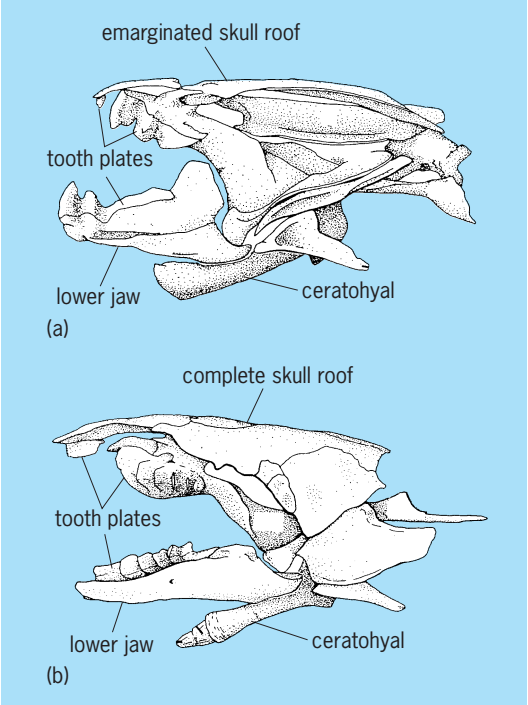

Fig. 1. Lateral views of the skulls of (*a*) African lungfish, *Protopterus aethiopicus*, and (*b*) Australian lungfish, *Neoceratodus forsteri*.

The family Ceratodontidae, represented by the single species *Neoceratodus forsteri* (**Fig. 2**), has a stout body and large scales, paddlelike paired fins, and tooth plates with flattened crushing surfaces. These fishes resemble typical Paleozoic lungfishes more closely than do the lepidosirenids. Adults reach more than 3 ft (1 m) in length. The single lung is poorly subdivided, and the gills are essential for gas exchange. During development, tooth plates are formed by the fusion of separate denticles. The diet includes plants, crustaceans, and soft-bodied invertebrates. Fossils referred to *N. forsteri* occur in the interior of Australia, but today the species is restricted to shallow, coastal rivers in Queensland, including the Burnett, Mary, and Brisbane rivers. *Neoceratodus forsteri* is completely protected by Australian law.

Some authorities ally lepidosirenids with Late Paleozoic genera such as *Gnathorhiza*, while others believe that the split between lepidosirenids and ceratodontids occurred more recently.

Phylogeny. Lungfishes show many characters seemingly intermediate to fishes and amphibians, and this has resulted in much controversy about the phylogenetic relationships of the group since the discovery of lepidosirenids in the 1830s. Most workers continue to regard rhipidistian fishes as more closely related to tetrapods than are lungfishes or coelacanths. William Bemis; Herbert Boschung

Fossils. Most Devonian dipnoans had two dorsal fins, a heterocercal tail, and cosmine (tissue composed of dentine, a pore-canal system, and a thin enamel layer) on scales and bones. Later forms lost the cosmine; their scales have a network of lines which separate each scale into plates. The caudal fin

Fig. 2. Australian lungfish (*Neoceratodus forsteri*). (*Photo © www.jjphoto.dk*)

became diphycercal and continuous with the dorsal and anal fins. The paired fins have a long, segmented axis with diverging radials, an archipterygium.

There are three types of dentition: rasping, deciduous denticles in the Early Devonian *Uranolophus* and the Late Devonian rhynchodipterids; smooth, crushing dental plates in a few Devonian forms; and crushing tooth plates with radiating ridges, which are the best-preserved fossil remains of the group, characteristic of most Devonian and all post-Devonian forms except *Conchopoma*.

Early dipnoans were marine; freshwater adaptations occur in Paleozoic and most post-Paleozoic dipnoans. Burrowing habits developed independently in Late Paleozoic gnathorhizids and extant lepidosirenids. Dipnoans were common worldwide from the Early Devonian to the end of the Triassic, but are rare since. They are restricted to the southern continents in the Cenozoic. *See* OSTEICHTHYES.

Hans-Peter Schultze

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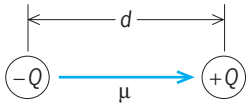
Dipole

Any object or system that is equally and oppositely charged at two points or poles, such as a magnet, a polar molecule, or an antenna element. The properties of a dipole are determined by its dipole moment, that is, the product of one of the charges and their separation. *See* DIPOLE MOMENT.

An electric dipole consists of two electric charges of equal magnitude but opposite polarity, separated by a short distance (see **illus.**); or more generally, a localized distribution of positive and negative electricity without net charge whose mean positions of positive and negative charge do not coincide.



Electric dipole with moment $\mu = Qd$.

Molecular dipoles which exist in the absence of an applied field are called permanent dipoles, while those produced by the action of a field are called induced dipoles. *See* POLAR MOLECULE.

The term magnetic dipole originally referred to the fact that a magnet has two poles and, because of these two poles, experiences a torque in a magnetic field if its axis is not along a magnetic flux line of the field. It is now generalized to include electric circuits which, because of the current, also experience torques in magnetic fields. *See* MAGNET.

The term dipole is also used to describe a commonly used element of radio-frequency and microwave antennas. An electric dipole whose moment oscillates sinusoidally radiates electromagnetic waves and is known as a hertzian dipole; it is of interest in developing the theory of electromagnetic radiation. For practical purposes, a half-wave dipole, consisting of two collinear conducting rods, fed at the center, whose combined length equals half the wavelength of the radiation to be transmitted or received, is often used as an antenna element, either by itself or in an array, or as a feed for a reflector. Other shapes of dipoles (using, for example, thick cylinders or double cones) may be used to achieve a greater bandwidth. *See* ANTENNA (ELECTROMAGNETISM); ELECTROMAGNETIC RADIATION; ELECTROMAGNETIC WAVE TRANSMISSION.

A. Earle Bailey

Dipole-dipole interaction

The interaction of two atoms, molecules, or nuclei by means of their electric or magnetic dipole moments. This is the first term of the multipole-multipole series of invariants. More precisely, the interaction occurs when one dipole is placed in the field of another dipole. The interaction energy depends on the strength and relative orientation of the two dipoles, as well as on the distance between the centers and the orientation of the radius vector connecting the centers with respect to the dipole vectors. The electric dipole-dipole interaction and magnetic dipole-dipole interaction must be distinguished.

Electric dipole-dipole interaction. The center of the negative charge distribution of a molecule may fail to coincide with its center of gravity, thus creating a dipole moment. An example is the water molecule. If such molecules are close together, there will be a

dipole-dipole interaction between them. Atoms do not have permanent dipole moments, but a dipole moment may be induced by the presence of another atom nearby; this is called the induced dipole-dipole interaction. Induced dipole-dipole forces between atoms and molecules are known by many different names: van der Waals forces, originally so named because they represent the attractive term in the van der Waals equation of state; London forces or dispersion forces, because F. London explained this interaction by using the picture of optical dispersion, a picture that was in turn based on the behavior of damped harmonic oscillators. The fact that the phases of the electrons in two different molecules are correlated in the dispersion formula led to the term dynamically induced dipole-dipole interaction. The proper way to treat the interaction is by means of a quantum-mechanical derivation using perturbation theory; the lowering of the energy resulting from the presence of one molecule in the neighborhood of the other is found to have the same features as the induced dipole-dipole interaction. These induced dipole-dipole forces are also responsible for cohesion and surface tension in liquids. They also act between unlike molecules, resulting in the adsorption of atoms on macroscopic objects. Moreover, using sophisticated experimental techniques, it can be shown that the same force exists between two macroscopic objects at very short distances. *See* ADSORPTION; CHEMICAL BONDING; COHESION (PHYSICS); DIELECTRIC MATERIALS; INTERMOLECULAR FORCES; SOLUTION; SURFACE TENSION; VAN DER WAALS EQUATION.

Magnetic dipole-dipole interaction. The magnetic dipole-dipole interaction is found both on a macroscopic and on a microscopic scale. The classical magnetic dipole is produced by a current loop. Its orientation is determined by the orientation of the surface element enclosed by the current loop. Another example is the permanent magnetic dipole of a compass needle; two of these within reasonable proximity of each other illustrate clearly the influence of the dipole-dipole interaction. In quantum mechanics, the magnetic moment is partially due to a current arising from the motion of the electrons in their orbits, and partially due to the intrinsic moment of the spin. The same interaction exists between nuclear spins. *See* ELECTRON SPIN; MAGNET; NUCLEAR MOMENTS.

Magnetic dipole-dipole forces are particularly important in low-temperature solid-state physics, the interaction between the spins of the ions in paramagnetic salts being a crucial element in the use of such salts as thermometers and as cooling substances. Usually the interaction is a mixture of exchange interaction and magnetic dipole-dipole interaction. In many magnetic rare-earth salts, the interaction is purely of the latter type. Since the dipole-dipole interaction is well understood, it is possible to make a calculation from first principles of the heat capacity and the susceptibility (deviation from the Curie law) of a salt. These determine its thermometric behavior. *See* ADIABATIC DEMAGNETIZATION; EXCHANGE INTERACTION; LOW-TEMPERATURE THERMOMETRY; MAGNETIC THERMOMETER.

The so-called high-temperature expansion of the partition function, used in the calculations of the interaction, leads to the need for lattice sums. The lattice sums for dipolar interactions are conditionally convergent; that is, the result of the summation depends on the arrangement of the terms. In terms of physical properties, this means that the magnetization depends on the shape of the sample. This is described by a shape-dependent demagnetization factor. *See* STATISTICAL MECHANICS.

In magnetic cooling and low-temperature thermometry, the dipole-dipole interaction provides the coupling between the spins, needed to make the system go to equilibrium. Its influence shows up at higher temperatures as a small deviation from the Curie law. *See* CURIE-WEISS LAW; DIPOLE; DIPOLE MOMENT.

Paul H. E. Meijer

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Dipole moment

A mathematical quantity characteristic of a dipole unit equal to the product of one of its charges times the vector distance separating the charges. The dipole moment μ associated with a distribution of electric charges q_i is given by Eq. (1), where $\mathbf{r}_i$ is

$$\mu = \sum_i q_i \mathbf{r}_i \tag{1}$$

the vector to the charge q_i. For systems with a net charge (for example, positive), the origin is taken at the mean position of the positive charges (and vice versa). Dipole moments have the dimensions coulomb-meters. Molecular dipole moments were previously expressed in debye units, where 1 debye $= 3.336 \times 10^{-30}$ C · m. *See* DIPOLE.

The electric potential Φ of a dipole at a long distance $\mathbf{R}$ from the dipole is given by Eq. (2), where θ

$$\Phi = \gamma \frac{|\mu| \cos\theta}{4\pi\epsilon_0 |\mathbf{R}|^2} \tag{2}$$

is the angle between μ and $\mathbf{R}$, ϵ_0 is the permittivity of vacuum, and γ is a geometrical factor. ($\epsilon_0 = 8.854 \times 10^{-12}$ farad/m and $\gamma = 1$ in SI units.)

The potential energy U of a dipole in a uniform electric field $\mathbf{E}$ is $U = -|\mu|\,|\mathbf{E}| \cos\phi$, where ϕ is the angle between μ and $\mathbf{E}$.

The induced dipole moment μ_i of a molecule may be expressed in terms of molecular parameters by the equation $\mu_i = \alpha \mathbf{E}_L$, where α is the polarizability and $\mathbf{E}_L$ is the local field strength acting at the molecular site. This relation permits the interpretation of the macroscopic polarization and hence

dipole moments in terms of molecular processes. *See* PERMITTIVITY; POLARIZATION OF DIELECTRICS.

Robert D. Waldron

Dipsacales

An order of flowering plants (angiosperms) in the asterid I group of the eudicotyledons consisting of around 6 families and almost 1000 species. Like many asterids of the asterid II group, Dipsacales have an inferior ovary and opposite leaves. They also have fewer stamens than petals (or petal lobes), and the flowers are often bilaterally symmetric or of irregular symmetry. Family limits have changed recently as a result of deoxyribonucleic acid (DNA) sequence studies, but the species content of the order is still similar to previous systems of classification. The flowers are typically arranged in heads, similar to those of their close relatives in Asterales and Apiales, but the ovary often contains more than one seed (reduced to a single seed in Dipsacaceae and Valerianaceae). Familiar members of Dipsacales include elderberry (*Sambucus*, Adoxaceae), honeysuckle (*Lonicera*, Caprifoliaceae), and teasel (*Dipsacus*, Dipsacaceae). *Viburnum* (Adoxaceae) and *Abelia* (Linnaeaceae) are commonly planted ornamental shrubs. *See* ASTERALES; MAGNOLIOPHYTA; PLANT KINGDOM.

Mark Chase

Diptera

An order of the class Insecta known as the true flies and so named because they possess only two wings. This characteristic, together with a pair of balancers or halteres, distinguishes flies from all other orders of the Insecta. The three names for members of the order, known commonly as flies, gnats, or midges, form a part of the common names of most families, genera, and species of the order. The term fly also forms a part of the compound names of the insects in many other orders, such as butterfly, Mayfly, and chalcid fly, but when used alone it is correctly applied only to the members of the Diptera. The forms most commonly known as maggots are actually dipterous larvae, and keds (Hippoboscidae) are parasitic forms of flies that have lost their wings. The Diptera are the most important group of insects considered in medical entomology. Many, especially the mosquitoes, are the vectors for numerous parasites and diseases.

In number of species, it is the third largest order; only the Coleoptera and Lepidoptera are larger. There are approximately 100,000 kinds of flies in the world, of which over 18,500 inhabit America north of Mexico.

The order is ubiquitous and is more widespread than any other of the insects. Diptera are found on every continent and most islands of the world, except for the coldest parts of the Arctic and Antarctic, and at the tops of extremely high mountain ranges. They occur in almost all available ecological niches. Specimens have been taken at altitudes of many thousands of feet by specially constructed traps attached to airplanes.

Taxonomy

In the early 1970s the classification of the Diptera underwent a general reappraisal. For convenience, the families of Diptera may be divided into the Nematocera and the Brachycera, and the latter group further divided into the Orthorrhapha and the Cyclorrhapha. Adults of the Nematocera are characterized by their slow flight and long antennae, the flagellum of which is divided into 10–65 segments. These are the more primitive Diptera, and the major groups are represented in fossil deposits as old as Jurassic and Cretaceous age.

The Brachycera are generally swift fliers, and their antennae are no longer than the head. The flagellum is usually reduced to a single segment with a threadlike process called an arista. In some families of Orthorrhapha, there is no arista and the flagellum may be superficially annulate. In the Orthorrhapha, the adult fly emerges from the pupa through a T-shaped anterior opening, while in the Cyclorrhapha it emerges from a circular opening at the anterior end of the puparium.

Following is a list of scientific and vernacular names of the more common families of flies.

- Nematocera
  - Trichoceridae (winter crane flies)
  - Ptychopteridae (phantom crane flies)
  - Tipulidae (crane flies)
  - Psychodidae (moth flies and sand flies)
  - Blepharoceridae (net-winged midges)
  - Chironomidae (midges)
  - Ceratopogonidae (biting midges)
  - Simuliidae (black flies)
  - Chaoboridae (phantom midges)
  - Culicidae (mosquitoes)
  - Dixidae (dixid midges)
  - Anisopodidae (wood gnats)
  - Bibionidae (march flies)
  - Scatopidae (scatopsid flies)
  - Mycetophilidae (fungus gnats)
  - Sciaridae (black fungus gnats)
  - Cecidomyiidae (gall midges)
- Brachycera—Orthorrhapha
  - Stratiomyidae (soldier flies)
  - Rhagionidae (snipe flies)
  - Tabanidae (horse flies and deer flies)
  - Bombyliidae (bee flies)
  - Asilidae (robber flies)
  - Empididae (dance flies)
  - Dolichopodidae (long-legged flies)
- Brachycera—Cyclorrhapha
  - Lonchopteridae (spear-winged flies)
  - Phoridae (hump-backed flies)
  - Platypezidae (flat-footed flies)
  - Syrphidae (syrphid flies)
  - Pipunculidae (big-headed flies)
  - Lonchaeidae (lonchaeid flies)
  - Lauxaniidae (lauxaniid flies)
  - Drosophilidae (pomace or fruit flies)

Brachycera—Cyclorrhapha (*cont.*)
- Ephydridae (shore flies)
- Anthomyiidae (anthomyiid flies)
- Muscidae (house flies)
- Calliphoridae (blow flies)
- Sarcophagidae (flesh flies)
- Oestridae (bot flies)
- Tachinidae (tachinid flies)
- Hippoboscidae (louse flies)
- Micropezidae (stilt-legged flies)
- Diopsidae (stalk-eyed flies)
- Sciomyzidae (marsh or snail-killing flies)
- Sepsidae (black scavenger flies)
- Heleomyzidae (heleomyzid flies)
- Sphaeoceridae (small dung flies)
- Agromyzidae (leaf-miner flies)
- Tethinidae (tethinid flies)
- Chloropidae (frit flies)
- Conopidae (thick-headed flies)
- Tephritidae (fruit flies)
- Piophilidae (skipper flies)

Morphology

Adult flies vary from somewhat less than 1 mm to over an inch in body length. The head is vertical, usually with three ocelli at the dorsal vertex and the mouthparts ventral. In some cases, however, the head is distinctly longer than high when viewed laterally. The compound eyes are usually large and prominent, and are either holoptic (touching at the top of the head) or dichoptic (separated at the vertex by part of the head capsule) (**Fig. 1**).

Mouthparts. The mouthparts are modified for either lapping or piercing, but never with the mandibles apposable and capable of chewing. The structure of the mouthparts varies greatly throughout the order so that their homologies are often rather difficult to determine. In the mosquitoes and some other Nematocera, the labrum, mandibles, hypopharynx, and maxillae are greatly elongated and form a piercing tube enclosed by the labium or labellum. Throughout the Orthorrhapha there tends to be a retention of mandibles for piercing, especially in those flies that require blood meals. In the Cyclorrhapha, on the other hand, mandibles tend to be reduced or lost entirely, and with few exceptions the work of the mouthparts is largely sucking, with the maxillae and labium taking over the large share of work. In these flies the labellum is often enlarged and flattened and provided with minute rasping spines and a specialized mechanism for taking up fluids. Since all adults flies imbibe fluids only, the presence of a pharyngeal "pump" to transfer fluids from the external substrate to the gut is characteristic.

Antennae. Each antenna typically has two basal segments and one or more additional segments called the flagellum. In the Nematocera the flagellum consists of a variable number of quite similar segments. In all remaining flies there are three distinct antennal segments. The third segment, representing the flagellum, is usually longer than the two basal ones and sometimes complex. It is either annulated or produced in various shapes, and often bears an arista which is bare or haired.

In the Cyclorrhapha, a small crescent-shaped sclerite, the lunule, is situated just above the antennae. The flies in which this sclerite is separated from the rest of the frontal portion by a suture have an internal distensible sac, or ptilinum, which can be pushed through this suture before the fly is fully developed. It is used to break open the end of the puparium to allow the fly to escape. Once hardened and fully adult, the fly loses all external trace of this sac.

Thorax. The prothorax and metathorax are small and closely united with the large mesothorax which contains the musculature for the single pair of wings. The legs are usually alike, except in species in which the prothoracic pair are raptorial and species in which the fore, mid, or hind legs of the males are modified by secondary sexual characters. The coxae are usually rather long, and there are from two to five tarsomeres. Claws are present in almost all species. Beneath each claw in many species there is a membranous pad, or pulvillus. In addition, some flies carry an empodium between each pair of pulvillae.

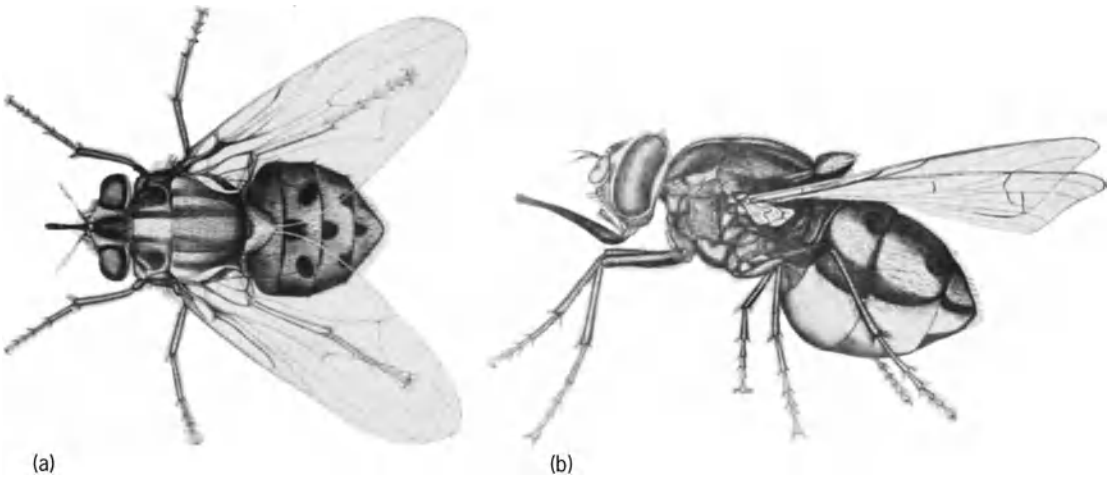

Fig. 1. Stable fly, *Stomoxys calcitrans*. (*a*) Adult female. (*b*) Female engorged with blood. (*USDA*)

This may be bristlelike or may take the form of membranous pad like the pulvillae in shape and size.

Wings. Only the first pair of wings, the mesothoracic pair, is developed. The second pair is reduced to a pair of club-shaped organs, or halteres, which serve as balancing organs in flight. They are present in most species, even when the mesothoracic wings are absent. In many species, especially those of the Nematocera, abundant scalelike setae clothe the veins and sometimes the margins of the wings. The wings of some of the higher Diptera bear yellow, brown, or black markings that aid in species identification. The venation of the primitive dipteran wing follows quite closely that of the primitive insect wing, except that vein radius (R) tends to be reduced, and only the important cross veins are present. In the higher Diptera the branches of some of the veins coalesce with others, thus further reducing the number of apparent veins present. In some Cyclorrhapha a pair of membranous lobes, the squamae or calypters, are found at the extreme base of the wing. Each of these occurs in pairs, the upper half fastened to the wing and the lower half to the thorax.

Setae. In certain families of Diptera, some of the setae covering the head and body are greatly enlarged and are used for identification. The location and relative size of these bristles are often characteristic of species, genera, or familes.

Eggs, larvae, and pupae. The eggs are usually elongated and taper at both ends. They are relatively smooth-surfaced, but the integument is often characteristically sculptured in designs that are visible only under high magnification (**Fig. 2**). The eggs of some aquatic species are provided with air sacs, of characteristic design, which aid in keeping them afloat. When laid, most eggs are whitish but darken gradually as the embryo grows within.

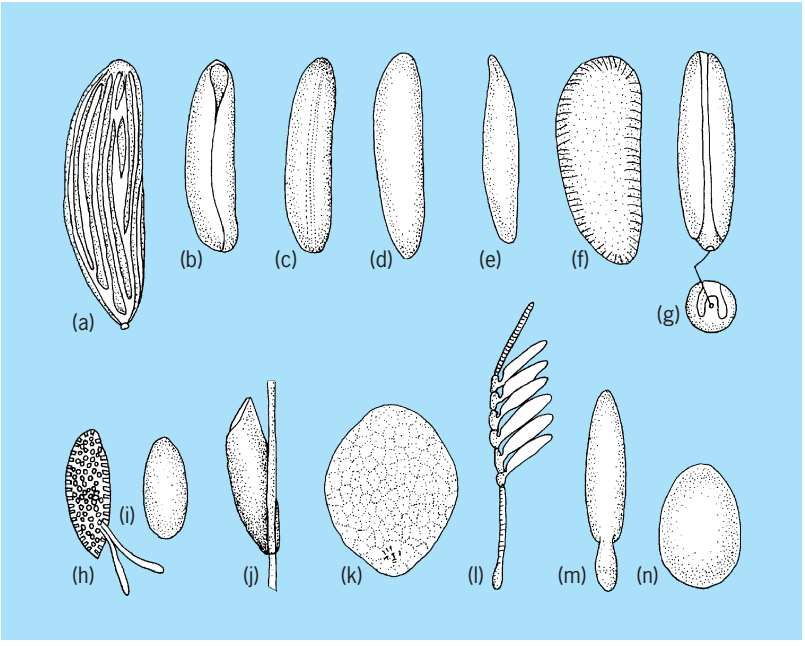

Fig. 2. Eggs of Diptera. (*a*) *Hippelates pusio*. (*b*) *Stomoxys calcitrans*. (c) *Cochliomyia macellaria*. (*d*) *Musca domestica*. (e) *Tabanus punctifer*. (*f*) *Cryptochaetum iceryae*. (*g*) *Cochliomyia hominovorax*. (*h*) *Drosophila melanogaster*. (*i*) *Tipula infuscata*. (*j*) *Gasterophilus nasalis*. (*k*) *Zenillia labatrix*. (*l*) *Lypoderma lineata* attached to a hair and (*m*) dorsal aspect of single egg. (*n*) *Simulium simile*. (*After E. O. Essig, College Entomology, Macmillan, 1942*)

Larvae, unlike adults, have elongate bodies. Mouthparts of this feeding stage are usually sclerotized structures. They are variously and, many times, intricately designed for rasping or chewing, but sometimes they are reduced or apparently entirely absent. The head of the larva of most lower Diptera is large, sclerotized, and complicated in internal structure, while in many species of higher Diptera it is reduced in size. The mouthparts are always well developed to enable the feeding stage to reach maturity readily. Although dipterous larvae never have true segmented appendages, anterior or posterior prolegs may be present. Nearly all types of spiracular arrangement are found in this order, but many larvae have prominent prothoracic and posterior spiracles with characteristic openings.

Pupae have appendages which are rather adherent to the body. In some pupae the developing adult is free, but in others it is encased in a hardened last larval skin resembling a seed or capsule and presenting no particular external features except spiracles and segmentation.

Life Cycle and Biology

The adult stage in the life history may be regarded as a reproductive and dispersal phase, during which these insects increase both in number and often in geographic range. Females select suitable oviposition sites and lay a variable number of eggs, after which they usually die. The eggs, after a period of incubation of varying length, give rise to larvae, which represent a growth phase. During this period, the insect does almost nothing but nourish itself, thus providing the tissues necessary for a later transformation into an adult. This growth period is divided into stages, or instars, each of which terminates with the molting of the larval skin to allow increase in size for the stage that follows. In the Diptera there are four larval instars. In most of the Orthorrhapha, especially those with aquatic immature forms, all four instars are active; but in the Cyclorrhapha, at the third larval molt, a hard puparium is formed inside which a quiescent fourth larval stage occurs. At the conclusion of the final larval instar, there is a dramatic reorganization in tissue structure initiated by the action of appropriate hormones on the so-called "imaginal buds," a process that eventually results in the formation of an adult insect. During this time, the insect is called a pupa because its external form is now altered and any external activity virtually ceases.

When the adult tissues are almost fully formed, the fly emerges from its pupal case and spends some time in drying, hardening, and expanding its wings, attaining color, and reaching sexual maturity. It is then ready to start its cycle once again (**Figs. 3** and **4**).

Adults. The adults, most of which are able to fly, represent the reproductive and dispersal phase of a fly's developmental cycle. After extricating itself from the puparium, the adult requires some period

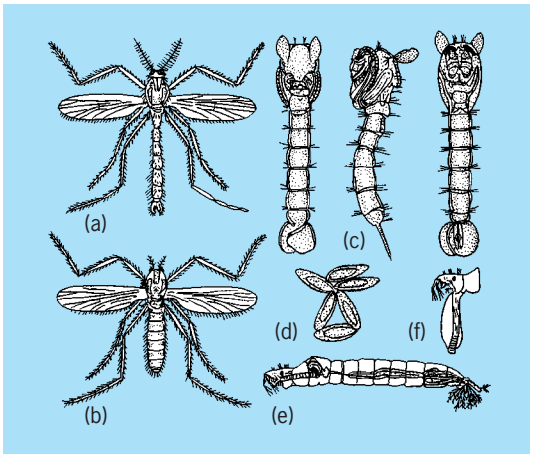

Fig. 3. Clear Lake gnat, *Chaoborus astictopus*. (*a*) Male; (*b*) female; (*c*) pupae; (*d*) eggs; (*e*) larva (phantom larva); (*f*) eversible pharyngeal sac extended from mouth. (*After E. O. Essig, College Entomology, Macmillan, 1942*)

of exposure to the air, after which it flies about to find food and to procreate.

Nutrition. Food requirements of females apparently differ from those of males because of their need for nutrient materials for developing and maturing eggs. In many of the nematocerous Diptera there is quite a different arrangement of mouthparts in the female, and nowhere is this more true than in the mosquitoes, of which most females require blood meals before they are able to develop eggs. Many female mosquitoes have decided preferences for certain hosts. Some few species, generally regarded as the most important disease vectors, are attracted principally to humans; others prefer a wide variety of mammals, and still others, birds.

The families of flies whose females require blood meals are scattered throughout the order but are concentrated mainly in two areas of the classification listing given earlier. The families Psychodidae, Ceratopogonidae, Culicidae, and Simuliidae have this habit. Females of the family Tabanidae also seek humans and other animals for blood. Among the calypterates, some Muscidae require blood, and many other species scattered among various families are attracted by exudates from the bodies of animals and humans.

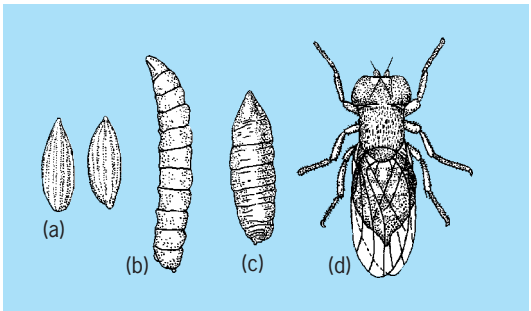

Fig. 4. Coachella eye gnat. *Hippelates pusio*. (*a*) Eggs; (*b*) larva; (*c*) pupa; and (*d*) adult. (*After E. O. Essig, College Entomology, Macmillan, 1942*)

The adults of a fairly large number of species are predators and feed on smaller anthropods. This habit is scattered throughout the order to such an extent that a listing of all families would be too lengthy. Whether such food materials are required for development of their eggs is not known with certainty.

Those flies whose larvae live and develop in plant materials produce adults with a great variety of food habits. It is known that adult Tephritidae of many species are especially attracted to exudates from scale insects, which are so numerous on many fruit trees that their honeydew covers large areas of foliage. It has been shown that certain components of this material are essential for the development of fruit fly eggs.

In the large majority of flies the food habits of adults have never been studied, and nothing is known except as the result of an occasional naturalist's observation.

Reproduction. Most female flies require a period of time between their emergence from the puparium and the time they are ready to lay eggs. The length of this period is sometimes characteristic of a species. Thus certain fruit flies are able to lay eggs 3 or 4 days after emergence while others require as long as 30 days. Females of many species will not accept males for copulation until a certain amount of time has expired; some species require fertilization after every egg-laying period, and some need fertilization only once during their lifetimes. Males usually emerge before females, presumably to assure that they will be sexually mature in time to fertilize the first eggs ready for oviposition.

Fertilization takes place in many ways throughout the order. Males of many of the nematocerous flies, and representatives of other families as well, form mating swarms. These are usually organized over an elevation in the surrounding terrain such as a mound of earth, bush, tree stump, or fence post. The swarm remains stationary over the object while the flies mill about in the swarm in every possible direction. Although such swarms are composed principally of males, some females are also present; the purpose of such swarms is to accomplish fertilization of those females.

The females of other species, especially many of the Tephritidae and perhaps of many other families, are fertilized at about the time they lay their eggs. Males of some of these species lie in wait on the surface of a fruit or some other substrate where they apparently know a female will alight to lay eggs. The male attacks just before, during, or immediately after the act of oviposition. Males have been seen to pull the ovipositor of a female from the pulp of a fruit to deposit the sperm.

Eggs. Study of fly eggs has been neglected by biologists, and most of their observations have been confined to the various ways in which females deposit them. Almost every conceivable kind of behavior accompanies egg laying. Some flies broadcast their small eggs while on the wing, spreading them about on the surrounding foliage. Eggs of some of these flies are eaten by caterpillars where they hatch

into larvae which feed within their hosts. Females of species whose larvae are parasitic cement their eggs on the surface of their host, or deposit them inside the bodies of the larvae, pupae, or adults that serve as their victims. Some females are able to deposit their eggs while their hosts are in full flight. Perhaps the most interesting habit is that of the female of *Dermatobia hominis*, which captures a passing mosquito upon which to fasten her eggs. As soon as this carrier has come to rest on an animal to feed, the *D. hominis* eggs quickly hatch and the larvae enter the mosquito's host.

Many flies whose larvae attack plant material have a sharply pointed, inflexible ovipositor to insert their eggs within the plant tissues, thus affording newly hatched larvae easy access to the succulent parts of the plant. Others without such an ovipositor fasten their eggs on the plant surface, and the newly hatched larvae immediately burrow into the plant tissue or seek a suitable place to gain entrance. The eggs of Sarcophagidae and some other flies hatch within the abdomen of the female and grow to various stages of development before escaping to continue a life of their own. In several genera of Cecidomyiidae, or gall midges, the reproductive organs develop prematurely with certain larvae and produce eggs which hatch internally so that only the new, or daughter, larvae escape from the body of the mother larva.

Larvae. It has been estimated that the immature stages of about one-half the known species of flies occur in association with water. Dipterous larvae are found in almost every conceivable habitat within this aquatic environment. Some occupy extreme situations such as intertidal wave-swept rocks on the seacoasts (Chironomidae), thermal inland waters as hot as 120°F or 49°C (Stratiomyidae), saturated brine pools (Ephydridae), and even natural petroleum seeping from the ground (Ephydridae). Adaptations such as these are unusual and occur only in a few species of the families mentioned. They are usually accounted for by special and unusual morphological adaptations.

Most aquatic larvae live in all kinds of standing water or in running water, from the smallest rivulet to the large slowly flowing rivers. The species that are found as aquatics in the ordinary sense, that is, swimming under and at the surface, are often provided with posterior spiracles especially adapted for obtaining atmospheric oxygen. These may be placed at the apex of a more or less elongated "air tube," such as that in many mosquito larvae, or on other specially adapted organs. They are almost always accompanied by a special arrangement of hairs and by a water-repellent coating so that water will not be "inhaled" at the surface. The larvae of some species may never come to the surface, but obtain oxygen by absorption through membranes in gills or through body surface membranes.

The remaining half of the order have larvae that inhabit nonaquatic habitats. Of these, a very large number of species are found in association with living plants. A large share of the Cecidomyiidae cause galls to form in plant tissue or live in galls produced by other species. Occasionally galls may be found underground in roots and stems, and often in aerial roots of epiphytes. More often they attack growing stems, leaves, fruits, and seeds, but no part of a plant can be mentioned that is not subject to a dipterous attack of one kind or another. Gall midges and certain tephritids characteristically make a large number of different kinds of galls. They are found as inquilines in galls made by other species and may even be found in masses of pitch made to exude from evergreens by their presence.

The fruit flies, Tephritidae, are so called because of their habit of feeding on ripe or nearly ripe fruit, or because they develop in the ovaries and seeds of those plants that do not bear succulent fruit. This occurs especially in tropical and subtropical areas of the world. Larvae of other species feed directly upon plant tissue, while certain Muscidae attack only roots. The Agromyzidae and certain other small acalypterates feed between layers of plant tissue in stems and leaves, causing mines that are often characteristic of the species making them.

Another large group of larvae, including those of another part of the Cecidomyiidae, are found in association with dead and decaying vegetation lying on the ground surface, and a number of tephritids and near relatives are strict saprophytes.

Many fly larvae are predaceous and are sometimes provided with specially adapted mouthparts for preying upon other species. True parasitism occurs here and there throughout the order. Larvae of some species of Tachinidae, Asilidaae, Bombyliidae, and others are parasitic in the bodies of a large number of insects. Sciomyzidae larvae feed principally on the soft bodies of snails; Pyrgotidae, on June beetles; Cryptochaetidae, on scale insects; and so forth.

Some of the more advanced flies of the calypterate group require the blood or other fluids of mammals for larval development. Most of these affect domestic livestock. Thus the primary screwworm causes wounds in the undamaged skin of large domestic animals, and the botflies (**Fig. 5**) have a complicated development during which their larvae travel into many parts of the bodies of their hosts.

The Congo floor maggot requires human or other animal blood and commonly lives in association with

Fig. 5. Botfly larvae, of a species of *Gasterophilus*, attached to inner lining of stomach of horse. (*Southwestern Biological Supply Co., Dallas, Texas*)

humans, feeding through the skin, but is not permanently attached. Larvae of the Hippoboscidae are carried full grown in the ovaries of the mother, while the mother, in turn, is a permanent resident of sheep and feeds on sheep's blood.

Pupae. As stated in the beginning section of this article, pupae may be of two kinds. In most, if not all, of the Orthorrhapha, all four larval instars are active feeding stages. They are either aquatic or phytophagous, and the skin of the last larval stage is cast free and lost at the formation of the pupa. Some pupae, especially of mosquitoes, are free-swimming just like the larvae, while others are quiescent forms. In the Cyclorrhapha, the skin of the third larval stage is retained and transformed into a barrel-shaped puparium inside which the fourth larval stage is quiescent and in which the pupa reaches full development. The various stages of this process have been described in detail for the apple maggot (*Rhagoletis pomonella*). Most cyclorrhaphous pupae are adapted for life underground or in other well protected parts of their environment, and larvae commonly burrow into the soil or these other locations to pupate.

No matter what the pupal environment, however, provision is always made for the escape of the adult. In many of the aquatic Diptera, pupation takes place at the air-water interface; mosquito pupae, for instance, simply float at the water surface until the adult is free. Adults of the Simuliidae rise to the water surface from pupae at the bottom surrounded by a bubble of air. Pupae of many species are provided with spines or hairs to help them move about.

Economic Importance

Diptera have probably more economic importance than any other insect order. This importance comes from the relatively few species that affect humans, domestic animals, and plants. The vast majority of forms have no direct importance although a number are beneficial in one way or another. Many flies visit flowers and are efficient plant pollinators. Adults may prey upon harmful arthropods or lay eggs in their bodies so that the developing larvae can use them as a source of food. Aquatic larvae and pupae under certain circumstances serve as abundant food for fish and other aquatic life, while the larvae of a number of terrestrial species act as scavengers, assisting bacteria to destroy all kinds of decaying animal and vegetable matter.

Human. Only a relatively few dipteran species cause severe economic loss to humans. Perhaps of most concern is the role played by them in disease transmission. About 70 species of *Anopheles* mosquitoes transmit an estimated 500,000 cases of malaria each year. This disease has serious debilitating effects and occurs on every continent and many islands of the world. Yellow fever, once a threat to completion of work on the Panama Canal, is transmitted principally by a single mosquito, *Aedes aegypti*. This disease is largely under control, yet the same virus is found in monkeys living in tropical rainforests of Central and South America and Africa, and is transmitted by *Aedes* and *Haemagogus* species to people who work out of doors in the jungle. Dengue, or breakbone fever, is a usually nonfatal disease of worldwide distribution that leaves its victims debilitated for several weeks. It is transmitted by *Aedes aegypti* and *A. albopictus*. Filariasis, primarily a disease of peoples of Africa, Asia, and the Pacific islands, is caused by a minute roundworm whose larvae are transmitted by a few species of *Anopheles*, *Mansonia*, *Culex*, and *Aedes*. Filiariasis affected a large number of American troops during World War II. Longstanding cases in natives of areas where the disease is endemic may result in an enlargement of the extremities called elephantiasis. With modern methods used by bacteriologists and virologists, a large number of virus diseases known to be transmitted by mosquitoes have been found in humans. Many of these, such as Sindbis virus of Egypt, may not produce clinical symptoms of disease, but others, such as the equine encephalitides, which include western, eastern, St. Louis, Japanese, and others, may be quickly fatal in humans. With the exception of the viruses, none of the diseases mentioned above are transmitted by any insects other than members of the order Diptera. *See* MALARIA; YELLOW FEVER.

A few less well-known diseases are transmitted by other flies. Black flies transmit *Onchocerca volvulus*, the causative agent of onchocerciasis, a disease affecting the eyes of natives of Central and parts of South America. Sandflies of the genus *Phlebotomus* transmit organisms that cause kala azar, Oriental sore, pappataci fever, and Oroya fever. Even today large parts of Africa remain underdeveloped because of the presence of sleeping sickness, a fatal disease transmitted by tsetse flies. Species of the genus *Chrysops* (Tabanidae) are instrumental in carrying tularemia, primarily a disease of rodents. *See* TULAREMIA.

Domestic animals. Domestic animals, hides, meat, and dairy products are affected by disease transmission or by direct attack by flies. Anthrax, tularemia, botulism, many virus diseases, and nagana, a form of sleeping sickness, are some of the diseases transmitted by members of the Diptera that take an annual toll amounting to millions of dollars.

Some flies wreak their damage by direct attack. The primary screwworm fly deposits eggs on hides of animals, and the larvae, upon hatching, burrow through the skin and into the flesh. The secondary screwworm gains entrance through holes, often infected, already present on the skin surface. *Hypoderma lineata* and *H. bovis* are botflies of special importance. Eggs of both species are laid on hairs of cattle, and the hatching larvae bore through the skin into the connective tissue. During their development they wander through the tissues of the animal, and when mature they escape through holes which they make along the spine of their host and drop to the ground to pupate. Horses are afflicted by a species of *Gasterophilus*. They lick the eggs from their bodies, and the larvae settle and dwell in their stomachs (Fig. 5), often in large numbers. Sheep are victims of the sheep botfly, whose larvae live in their nasal passages and sinuses. All of these insects

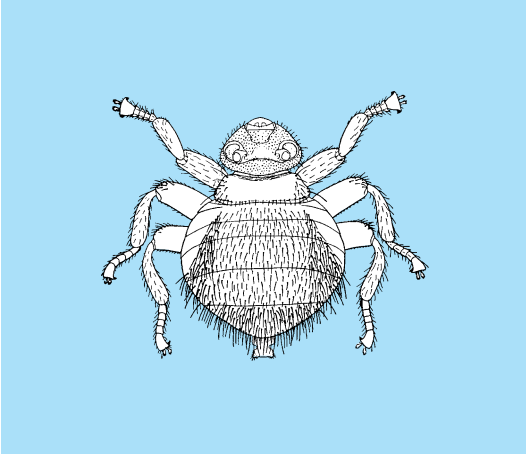

Fig. 6. Bee louse, *Braula coeca*. (*After E. O. Essig, College Entomology, Macmillan, 1942*)

affect the meat, milk, and wool production of animals through debilitation and irritation, and hides may be rendered completely useless by their entrance and exit holes. An interesting dipteran is the bee louse, which is a wingless, ectoparasite of the honeybee (**Fig. 6**).

Many species of flies are attracted by unsanitary conditions about garbage cans and rendering plants, by undressed wounds, and by seepages from infected eyes and mucous membranes of living animals. By coming into contact with these products, flies can mechanically carry on their feet and body hairs the organisms that cause typhoid, dysentery, diarrhea, cholera, yaws, and trachoma, in addition to certain parasitic worms. *See* MYIASIS.

Plants. Economic losses as a result of damage by flies to crops are perhaps not as great as those caused by other insects, yet their presence has necessitated the expenditure of large sums of money for control. Among these, perhaps the most important are members of the family Tephritidae, the larvae of which feed upon and ruin the succulent flesh and seeds of their hosts. The most important of these are the European and American cherry fruit flies; their relatives in the genus *Rhagoletis* that attack walnuts in North America; six or seven species of the genus *Anastrepha* which attack many kinds of fruit in the New World; the Mediterranean fruit fly which is a limiting factor in the production of many fruits, especially citrus, in many parts of Africa, Central and South America, and the Mediterranean Basin; and several species of the genus *Dacus*, which attack olives, citrus fruits, many kinds of vegetables, and other edible plants and plant parts in Europe, Asia, and Pacific islands.

Larvae of some flies mine the leaves of ornamentals thereby defacing them, reducing their growth potentials, and affecting their production and sale by nurseries. The Hessian fly, a gall midge, is a serious pest of wheat and other grains grown for human consumption in Europe, northern Asia and North America, and causes untold losses to growers in those areas. Several other species of Itonididae affect the growth of rice and other plants grown for human consumption.

The larvae of some species of Muscidae are known as root maggots and burrow into the underground parts of plants with considerable loss to growers of truck crops over the world. *See* ARTHROPODA; ENTOMOLOGY, ECONOMIC; INSECTA; PLANT PATHOLOGY.

Raymond J. Gagne

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Direct-coupled amplifier

A device for amplifying signals with direct-current components. There are many different situations where it is necessary to amplify signals having a frequency spectrum which extends to zero. Some typical examples are amplifiers in electronic differential analyzers (analog computers), certain types of feedback control systems, medical instruments such as the electrocardiograph, and instrumentation amplifiers. Amplifiers which have capacitor coupling between stages are not usable in these cases, because the gain at zero frequency is zero. Therefore, a special form of amplifier, called a dc (direct-current) or direct-coupled amplifier, is necessary. These amplifiers will also amplify alternating-current (ac) signals. *See* AMPLIFIER; ANALOG COMPUTER; BIOMEDICAL ENGINEERING; CONTROL SYSTEMS; INSTRUMENTATION AMPLIFIER.

Some type of coupling circuit must be used between successive amplifier stages to prevent the relatively large supply voltage of one stage from appearing at the input of the following stage. These circuits must pass dc signals with the least possible amount of attenuation.

Interstage direct-coupling in transistor dc amplifiers must be implemented with special care. The use of both *npn* and *pnp* transistors is a possible solution. However, the *pnp* transistors available in monolithic form have relatively poor current-gain and frequency-response characteristics. If a dc amplifier is formed by a cascade of *npn* stages, there is a positive dc level buildup toward the positive supply voltage. This voltage buildup limits the linearity and amplitude of the available output swing. The problem can be overcome by using a level-shift stage between each stage to shift the output dc level toward the negative supply with minimum attenuation of the amplified signal. Practical dc level-shift stages suitable for monolithic circuit applications can use Zener diodes, (**Fig. 1***a*), a series of diodes (Fig. 1*b*), or a V_{BE}, multiplier circuit (Fig. 1*c*). For example, the Zener diode level-shift circuit (Fig. 1*a*) provides

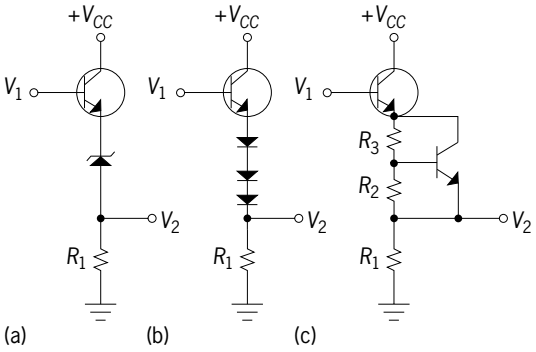

Fig. 1. Level-shift stages using (*a*) Zener diodes, (*b*) series of diodes, and (*c*) V_{BE} multiplier circuit. V_1 = input voltage, V_2 = output voltage.

a means of shifting the dc level between the input and the output by an amount equal to the sum of the base-to-emitter voltage and the Zener voltage. If the bulk resistance of the Zener diode is negligible compared to the resistance between the output terminal and ground (R_1), the voltage gain for the stage will be approximately unity. In the V_{BE} multiplier circuit (Fig. 1*c*), the output (V_2) is related to the input (V_1) by the equation below, where V_{BE} is the base-

$$V_1 = V_2 + V_{BE}\left(1 + \frac{R_3}{R_2}\right)$$

to-emitter voltage and R_2 and R_3 are two resistances in the circuit. *See* VOLTAGE-MULTIPLIER CIRCUIT.

Differential dc amplifier. It is generally recognized that the differential amplifier (**Fig. 2**) is the most stable dc amplifier circuit available. This is true because in this circuit the performance depends on the difference of the device parameters, and transistors can be manufactured using the planar epitaxial technique with very close matching of their parameters.

Carrier amplifier. A method of amplifying dc (or slowly varying) signals by means of ac amplifiers is to modulate a carrier signal by the signal to be amplified, amplifying the modulated signal, and demodulating at the output. (In some applications, such as instrument servomechanisms, output in the form of a modulated carrier is required and no demodulation is necessary.)

An analysis of an actual circuit would show that in order to have an output free from harmonics introduced by the modulation, the output low-pass filter cutoff frequency must be small compared to the modulation frequency. This limits the bandwidth of the input signal to a small fraction of the bandwidth of signals which can be amplified by the types previously discussed. This is a disadvantage in most cases, but there are cases, such as the output voltage from a thermocouple, where the required bandwidth is small. Various types of choppers (bipolar transistors, field-effect transistors) are used as modulators. An additional disadvantage is that the chopper must be carefully designed to reduce to a minimum the noise and the offset voltage and leakage current of a transistor. However, this amplifier provides stable amplification of the input signal, and it is used in many industrial recording instruments. *See* MODULATION; VOLTAGE AMPLIFIER.

Chopper-stabilized amplifiers. The offset voltage of matched transistor pairs of differential amplifiers can be a source of serious problems in precision analog dc amplifier applications. Typically the offset voltage of matched metal oxide semiconductor (MOS) transistor pairs can be reduced to within ±20 mV by careful processing. However, even this low offset voltage in many applications is unacceptable. It is possible to reduce the effective input offset voltage to below ±1 mV by using chopper-stabilized amplifiers employing offset-nulling or auto-zero techniques. These techniques are essentially sampled-data methods and are based on the concept of measuring periodically the offset voltage and subsequently storing it as a voltage across a holding capacitor and then subtracting it from the signal plus the offset. Christos C. Halkias

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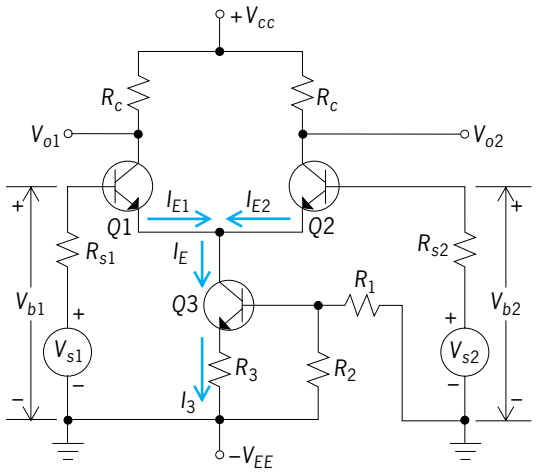


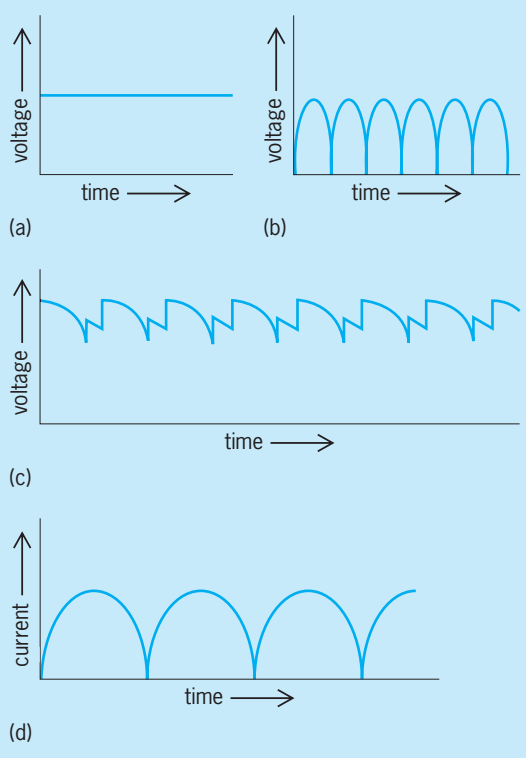
Fig. 2. Differential dc amplifier with constant-current stage in the emitter circuit.

Direct current

Electric current which flows in one direction only through a circuit or equipment. The associated direct voltages, in contrast to alternating voltages, are of unchanging polarity. Direct current corresponds to a drift or displacement of electric charge in one unvarying direction around the closed loop or loops of an electric circuit. Direct currents and voltages may be of constant magnitude or may vary with time.

Batteries and rotating generators produce direct voltages of nominally constant magnitude (**illus. *a***). Direct voltages of time-varying magnitude are produced by rectifiers, which convert alternating voltage to pulsating direct voltage (illus. *b–d*). *See* BATTERY; GENERATOR; RECTIFIER.

Direct current is used extensively to power adjustable-speed motor drives in industry and in transportation. Very large amounts of power are used in electrochemical processes for the refining and plating of metals and for the production



Typical direct currents and voltages. (*a*) Battery voltage; (*b*) full-wave rectified voltage; (*c*) dc voltage of the rectifier station of a high-voltage transmission link (*after E. W. Kimbark, Direct Current Transmission, vol. 1, Wiley-Interscience, 1971*). (*d*) Direct current in a resistance supplied by a full-wave rectifier.

of numerous basic chemicals. *See* DIRECT-CURRENT MOTOR; ELECTROCHEMICAL PROCESS; ELECTROPLATING OF METALS.

Direct current ordinarily is not widely distributed for general use by electric utility customers. Instead, direct-current (dc) power is obtained at the site where it is needed by the rectification of commercially available alternating current (ac) power to dc power. Solid-state rectifiers ordinarily are employed to supply dc equipment from ac supply lines. Rectifier dc supplies range from tiny devices in small electronic equipment to high-voltage dc transmission links of several thousand megawatts capacity. *See* ELECTRIC POWER SYSTEMS; SEMICONDUCTOR RECTIFIER.

Many high-voltage dc transmission systems have been constructed throughout the world. Very large amounts of power, generated as ac and ultimately used as ac, are transmitted as dc power. Rectifiers supply the sending end of the dc link; inverters then supply the receiving-end ac power system from the link. For long transmission distances, the ac-dc-ac system often is more economical than ac transmission, especially in the case of underground or underwater cable circuits. *See* ALTERNATING CURRENT; ELECTRIC CURRENT; TRANSMISSION LINES. D. D. Robb

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Direct-current circuit theory

An analysis of relationships within a dc circuit. Any combination of direct-current (dc) voltage or current sources, such as generators and batteries, in conjunction with transmission lines, resistors, inductors, capacitors, and power converters such as motors is termed a dc circuit. Historically the dc circuit was the first to be studied and analyzed mathematically. *See* CIRCUIT (ELECTRICITY).

Classification. Circuits may be identified and classified into simple series and parallel circuits. More complicated circuits may be developed as combinations of these basic circuits.

Series circuit. A series circuit is illustrated in **Fig. 1**. It consists of a battery of E volts and three resistors of resistances R_1, R_2, and R_3, respectively. The conventional current flows from the positive battery terminal through the external circuit and back to the negative battery terminal. It passes through each resistor in turn; therefore the resistors are in series with the battery.

Parallel circuit. The parallel circuit, shown in **Fig. 2**, consists of a battery paralleled by three resistors. In this case the current leaving the positive terminal of the battery splits into three components, one component flowing through each resistor, then recombining into the original current and returning to the negative terminal of the battery.

Physical laws of circuit analysis. The operation of the basic series and parallel circuits must obey certain fundamental laws of physics. These laws are referred to as Ohm's law and Kirchhoff's laws in honor of their originators.

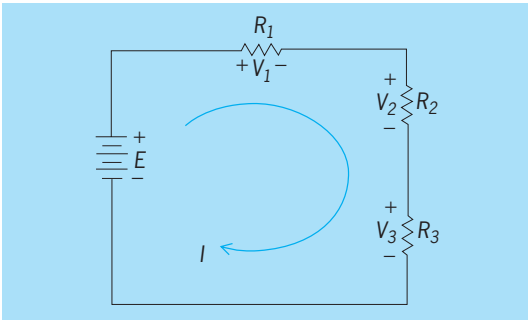

Fig. 1. Simple series circuit.

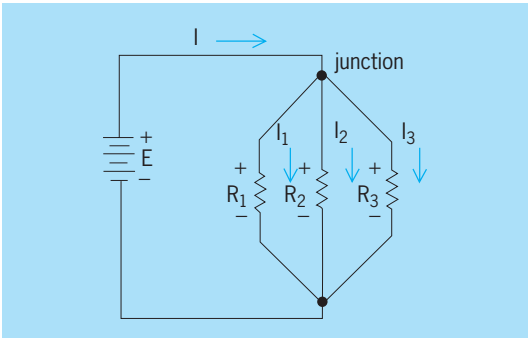

Fig. 2. Simple parallel circuit.

Voltage drops. When an electric current flows through a resistor, a voltage drop appears across the resistor, the polarity being such that the voltage is positive at the end where the conventional current enters the resistor. This voltage drop is directly proportional to the product of the current in amperes and the resistance in ohms. This is Ohm's law, expressed mathematically in Eq. (1). Thus in Fig. 1 the

$$V = IR \tag{1}$$

drop across R_1 is V_1 and has the polarity shown. *See* OHM'S LAW.

Summation of voltages. The algebraic sum of all voltage sources (rises) and voltage drops must add up to zero around any closed path in any circuit. This is Kirchhoff's first law. In Fig. 1, the sum of the voltages about this closed circuit is as given by Eqs. (2),

$$E - V_1 - V_2 - V_3 = 0 \tag{2a}$$

$$E = V_1 + V_2 + V_3 \tag{2b}$$

where the minus signs indicate a voltage drop. Written in terms of current and resistance, this becomes Eq. (3). From this results the important conclusion

$$E = I(R_1 + R_2 + R_3) = IR_{\text{total}} \tag{3}$$

that resistors in series may be added to obtain the equivalent total resistance, as shown in Eq. (4).

$$R_{\text{total}} = R_1 + R_2 + R_3 + \cdots \tag{4}$$

See KIRCHHOFF'S LAWS OF ELECTRIC CIRCUITS.

Summation of currents. The algebraic sum of all currents flowing into a circuit junction must equal the algebraic sum of all currents flowing out of the junction. In the circuit shown in Fig. 2, the current flowing into the junction is I amperes while that flowing out is the sum of I_1 plus I_2 plus I_3. Therefore the currents are related by Eq. (5). This is Kirchhoff's

$$I = I_1 + I_2 + I_3 \tag{5}$$

second law. In this case the same voltage appears across each resistor. If the currents are expressed in terms of this voltage and values of the individual resistors by means of Ohm's law, Eq. (5) becomes Eq. (6). The equivalent resistance R_{eq} that can re-

$$I = \frac{E}{R_1} + \frac{E}{R_2} + \frac{E}{R_3} = \frac{E}{R_{\text{eq}}} \tag{6}$$

place the resistances in parallel can be obtained by solving Eq. (6) in the form of Eq. (7). Therefore, re-

$$R_{\text{eq}} = \left(\frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3} + \cdots\right)^{-1} \tag{7}$$

sistances in parallel are added by computing the corresponding conductances (reciprocal of resistance) and adding to obtain the equivalent conductance. The reciprocal of the equivalent conductance is the equivalent resistance of the parallel combination. *See* CONDUCTANCE.

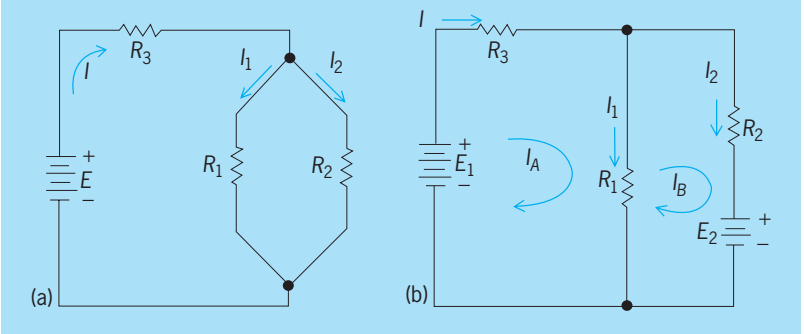

Fig. 3. Series-parallel circuits. (*a*) Single source. (*b*) Multiple sources.

Sources. Sources such as batteries and generators may be connected in series or parallel. Series connections serve to increase the voltage; the net voltage is the algebraic sum of the individual source voltages. *See* BATTERY; GENERATOR.

Sources in parallel provide the practical function of increasing the net current rating over the rating of the individual sources; the net current rating is the sum of the individual current ratings.

Series-parallel circuits. More complicated circuits are nothing more than combinations of simple series and parallel circuits as illustrated in **Fig. 3**.

Single source. Circuits that contain only a single source are readily reduced to a simple series circuit. In the circuit of Fig. 3*a* the parallel combination of R_1 and R_2 is computed and used to replace the parallel combination. The resultant circuit is now a simple series circuit consisting of R_3 and R_{eq} readily be solved for the series current if the voltage is known.

Multiple sources. Circuits that contain two or more sources located in various branches cannot be reduced to a simple series circuit (Fig. 3*b*). The three basic laws of circuit theory still hold and may be directly applied to provide a simultaneous solution to the loop currents I_A and I_B flowing in each basic series circuit present in the overall network.

In this example the summations of voltages around the individual series circuits or loops are given by Eqs. (8) and (9), which may be solved for the mathe-

$$E_1 = (R_1 + R_3)I_A - R_1 I_B \tag{8}$$

$$-E_2 = -R_1 I_A + (R_1 + R_2)I_B \tag{9}$$

matical loop currents I_A and I_B. These loop currents can in turn be identified by reference to the circuit where I_A is identical to I, and therefore I_B, as stated by Eq. (10).

$$I_A - I_B = I_1 \tag{10}$$

This method may be used to solve any complicated combination of simple circuits. Other methods are also available to the circuit analyst. *See* NETWORK THEORY.

Power. The electric power converted to heat in any resistance is equal to the product of the voltage drop across the resistance times the current through the resistance, as stated by Eq. (11*a*). By means of Ohm's law, Eq. (11*a*) may also be written as

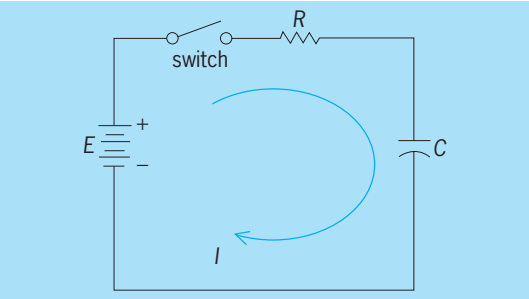

Fig. 4. A series *RC* circuit.

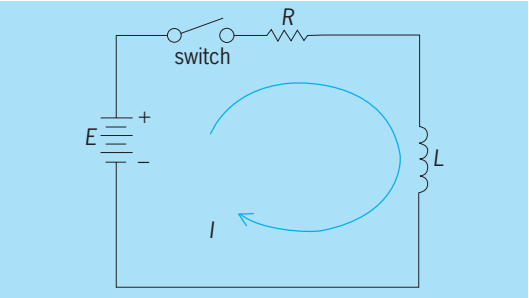

Fig. 5. A series *RL* circuit.

Eq. (11*b*). The total power dissipated in a circuit is

$$P = VI \tag{11a}$$

$$P = VI = \frac{V^2}{R} = I^2R \tag{11b}$$

the arithmetic sum of the power dissipated in each resistance.

Circuit response. In the circuits mentioned thus far, the circuit responds in an identical manner from the moment the circuit is excited (switches closed) through any extended period of time. This is not true of circuits typified by those of **Figs. 4** and **5**.

For instance, when the switch of Fig. 4 is first closed, a momentary current limited only by resistance R flows. As time passes, the capacitor, with capacitance C, charges and the voltage across it increases, eventually reaching a value equal to the applied voltage, at which time all flow of current ceases. The circuit current, given by Eq. (12), is in

$$i = \frac{E}{R}\epsilon^{-t/RC} \tag{12}$$

amperes. The product RC is known as the time constant of the circuit. The energy W in joules stored in a capacitance at any time is given by Eq. (13).

$$W = \frac{CE^2}{2} \tag{13}$$

See TIME CONSTANT.

For the circuit of Fig. 5, the initial current upon closing the switch is zero, since any attempt to cause a rate of change of current through the coil whose inductance is L induces a counter emf across the coil or inductor. Eventually this counter emf disappears and a steady-state current E/R flows indefinitely in the circuit.

The current at any time after closing the switch is given by Eq. (14) in amperes. The factor R/L is the

$$i = \frac{E}{R}(1 - \epsilon^{-Rt/L}) \tag{14}$$

time constant of the circuit. The energy W stored in an inductance at any time is given by Eq. (15).

$$W = \frac{LI^2}{2} \tag{15}$$

For a complete discussion of transient phenomena *see* ELECTRIC TRANSIENT

Robert L. Ramey

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Direct-current generator

A rotating electric machine which delivers a unidirectional voltage and current. An armature winding mounted on the rotor supplies the electric power output. One or more field windings mounted on the stator establish the magnetic flux in the air gap. A voltage is induced in the armature coils as a result of the relative motion between the coils and the air gap flux. Faraday's law states that the voltage induced is determined by the time rate of change of flux linkages with the winding. Since these induced voltages are alternating, a means of rectification is necessary to deliver direct current at the generator terminals. Rectification is done by a commutator mounted on the rotor shaft. *See* COMMUTATION; WINDINGS IN ELECTRIC MACHINERY.

Carbon brushes, insulated from the machine frame and secured in brush holders, transfer the armature current from the rotating commutator to the external circuit. Brushes are held against the commutator under a pressure of 2–2.5 lb/in.^{2} (14–17 kilopascals). Armature current passes from the brush to brush holder through a flexible copper lead. In multipolar machines all positive brush studs are connected together, as are all negative studs, to form the positive and negative generator terminals. In most dc generators the number of brush studs is the same as the number of main poles. In modern machines brushes are located in the neutral position where the voltage induced in a short-circuited coil by the main pole flux is zero. The brushes continuously pick up a fixed, instantaneous value of the voltage generated in the armature winding. *See* COMMUTATOR.

The generated voltage is dependent upon speed n in revolutions per minute, number of poles p, flux per pole Φ in webers, number of armature conductors z, and the number of armature paths a. The

equation for the average voltage generated is

$$E_g = \frac{np\Phi z}{60a} \text{ volts}$$

The field windings of dc generators require a direct current to produce a magnetomotive force (mmf) and establish a magnetic flux path across the air gap and through the armature. Generators are classified as series, shunt, compound, or separately excited, according to the manner of supplying the field excitation current. In the separately excited generator, the field winding is connected to an independent external source. Using the armature as a source of supply for the field current, dc generators are also capable of self-excitation. Residual magnetism in the field poles is necessary for self-excitation. Series, shunt, and compound-wound generators are self-excited, and each produces different voltage characteristics.

When operated under load, the terminal voltage changes with change of load because of armature resistance drop, change in field current, and armature reaction. Interpoles and compensating, or poleface, windings are employed in modern generators in order to improve commutation and to compensate for armature reaction. *See* ARMATURE REACTION.

Series generator. The armature winding and field winding of this generator are connected in series, as shown in **Fig. 1**. Terminals T_1 and T_2 are connected to the external load. The field mmf aids the residual magnetism in the poles, permitting the generator to build up voltage. The field winding is wound on the pole core with a comparatively few turns of wire of large cross section capable of carrying rated load current. The magnetic flux and consequently the generated emf and terminal voltage increase with increasing load current. **Figure 2** shows the external characteristic or variation of terminal voltage with load current at constant speed. Series generators are suitable for special purposes only, such as a booster in a constant voltage system, and are therefore seldom employed.

Shunt generator. The field winding of a shunt generator is connected in parallel with the armature winding, as shown in **Fig. 3**. The armature supplies both load current I_t and field current I_f. The field current is 1–5% of the rated armature current I_a, the higher value applying to small machines. The field winding resistance is fairly high since the field consists of many turns of small-cross-section wire. For voltage buildup the total field-circuit resistance must be below a critical value; above this value the generator voltage cannot build up. The no-load voltage to which the generator builds up is varied by means of a rheostat in the field circuit. The external voltage characteristic (Fig. 2) shows a reduction of voltage with increases in load current, but voltage regulation is fairly good in large generators. The output voltage may be kept constant for varying load current conditions by manual or automatic control of the rheostat in the field circuit. A shunt generator will not maintain a large current in a short circuit in the external circuit, since the field current at short circuit is zero.

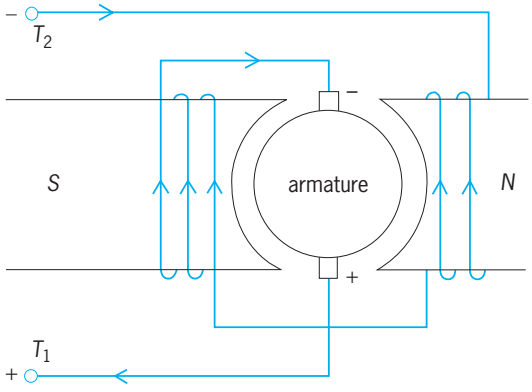

Fig. 1. Series generator.

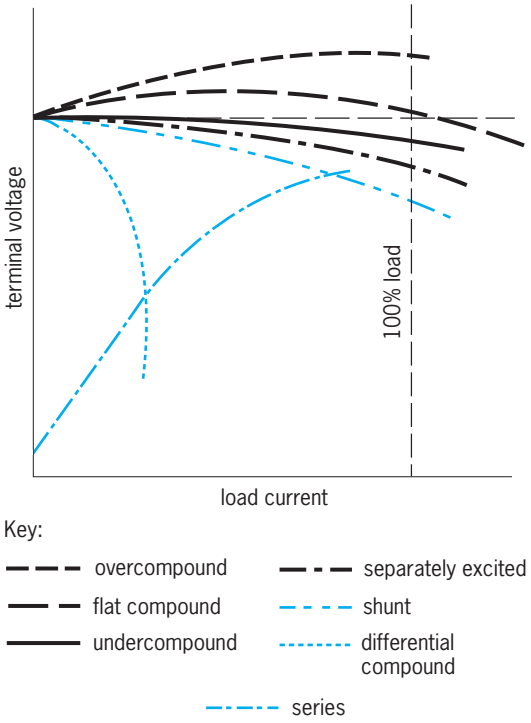

Fig. 2. External characteristics of dc generators.;

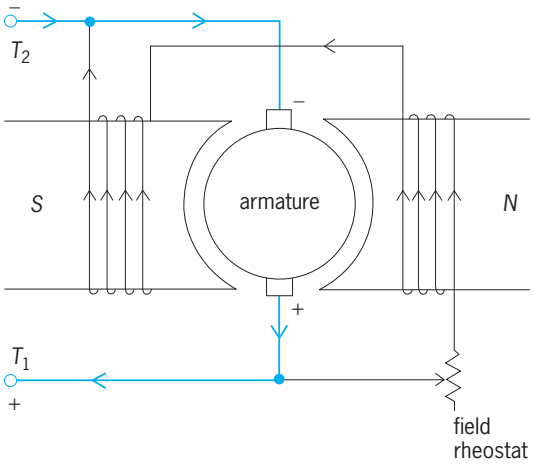

Fig. 3. Shunt generator.

The shunt generator is suitable for fairly constant voltage applications, such as an exciter for ac generator fields, battery charging, and for electrolytic work requiring low-voltage and high-current capacity. Prior to the use of the alternating-current generator and solid-stage rectifying devices in automobiles, a shunt generator, in conjunction with automatic regulating devices, was used to charge the battery and supply power to the electrical system. Shunt-wound generators are well adapted to stable operation in parallel.

Compound generator. This generator has both a series field winding and a shunt field winding. Both windings are on the main poles with the series winding on the outside. The shunt winding furnishes the major part of the mmf. The series winding produces a variable mmf, dependent upon the load current, and offers a means of compensating for voltage drop. **Figure 4** shows a cumulative-compound connection with series and shunt fields aiding. A diverter resistance across the series field is used to adjust the series field mmf and vary the degree of compounding. By proper adjustment a nearly flat output voltage characteristic is possible. Cumulative-compound generators are overcompounded, flat-compounded, or undercompounded, as shown by the external characteristics in Fig. 2. The shunt winding is connected across the armature (short-shunt connection) or across the output terminals (long-shunt connection). Figure 4 shows the long-shunt connection.

Voltage is controllable over a limited range by a rheostat in the shunt field circuit. Compound generators are used for applications requiring constant voltage, such as lighting and motor loads. Generators used for this service are rated at 125 or 250 V and are flat or overcompounded to give a regulation of about 2%. An important application is in steel mills which have a large dc motor load. Cumulative-compound generators are capable of stable operation in parallel if the series fields are connected in parallel on the armature side by means of a low-resistance cable or bus referred to as an equalizer bus. By paralleling the series fields of all compound generators, the equalizer bus maintains the same voltage across all series fields.

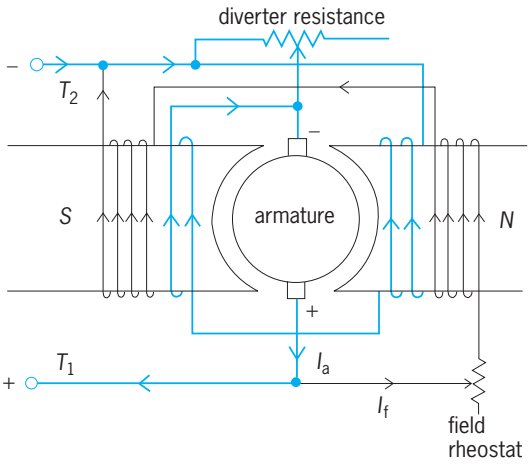

Fig. 4. Cumulative compound generator. The long-shunt connection is seen in this example.

In the differentially compounded generator the series field is connected to oppose the shunt field mmf. Increasing load current causes a large voltage drop due to the demagnetizing effect of the series field. The differentially compounded generator has only a few applications, such as arc-welding generators and special generators for electrically operated shovels.

Separately excited generator. The field winding of this type of generator is connected to an independent dc source. The field winding is similar to that in the shunt generator. Separately excited generators are among the most common of dc generators, for they permit stable operation over a very wide range of output voltages. The slightly drooping voltage characteristic (Fig. 2) may be corrected by rheostatic control in the field circuit. Applications are found in special regulating sets and in laboratory and commercial test sets.

Special types. Besides the common dc generators discussed in this article there are a number of special types, including the homopolar, third-brush, diverter-pole, and Rosenberg generators. For discussion of the Amplidyne, Regulex, and Rototrol *see* DIRECT-CURRENT MOTOR

Commutator ripple. The voltage at the brushes of dc generators is not absolutely constant. A slight high-frequency variation exists, which is superimposed upon the average voltage output. This is called commutator ripple and is caused by the cyclic change in the number of commutator bars contacting the brushes as the machine rotates. The ripple decreases as the number of commutator bars is increased and is usually ignored. In servomechanisms employing a dc tachometer for velocity feedback, the ripple frequency is kept as high as possible. *See* ELECTRIC ROTATING MACHINERY; GENERATOR.

Robert T. Weil, Jr.

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Direct-current motor

An electric rotating machine energized by direct current and used to convert electric energy to mechanical energy. It is characterized by its relative ease of speed control and, in the case of the series-connected motor, by an ability to produce large torque under load without taking excessive current. Output of this motor is given in horsepower, the unit of mechanical power. Normal full-load values of voltage, current, and speed are generally given.

Direct-current motors are manufactured in several horsepower-rating classifications: (1) subfractional, approximately 1–35 millihorsepower (mhp) or 0.75–25 W; (2) fractional, $^1/_{40}$–1 horsepower (hp)

or 20–750 W; and (3) integral, $^1/_2$ to several hundred horsepower or 0.4 to several hundred kilowatts.

The standard line voltages applied to dc motors are 6, 12, 27, 32, 115, 230, and 550 V. Occasionally they reach higher values.

Normal full-load speeds are 850, 1140, 1725, and 3450 rpm. Variable-speed motors may have limiting rpm values stated.

Protection of the motor is afforded by several types of enclosures, such as splash-proof, drip-proof, dust-explosion-proof, dust-ignition-proof, and immersion-proof enclosures. Some motors are totally enclosed.

The principal parts of a dc motor are the frame, the armature, the field poles and windings, and the commutator and brush assemblies (**Fig. 1**). The frame consists of a steel yoke of open cylindrical shape mounted on a base. Salient field poles of sheet-steel laminations are fastened to the inside of the yoke. Field windings placed on the field poles are interconnected to form the complete field winding circuit. The armature consists of a cylindrical core of sheet-steel disks punched with peripheral slots, air ducts, and shaft hole. These punchings are aligned on a steel shaft on which is also mounted the commutator. The commutator, made of hard-drawn copper segments, is insulated from the shaft. The segments are insulated from each other by mica. Stationary carbon brushes in brush holders make contact with the segments of the commutator. Copper conductors which are placed in the insulated armature slots are interconnected to form a reentrant lap or wave style of winding. *See* COMMUTATION; WINDINGS IN ELECTRIC MACHINERY.

Principles. Rotation of a dc motor is produced by an electromagnetic force exerted upon current-carrying conductors in a magnetic field. For basic principles of motor action *see* MOTOR

In **Fig. 2**, forces act on conductors on the left path of the armature to produce clockwise rotation. Those conductors on the right path, whose current direction is reversed, also will have forces to produce clockwise rotation. The action of the commutator allows the current direction to be reversed as a conductor passes a brush.

The net force from all conductors acting over an average radial length to the shaft center produces a torque T given by Eq. (1), where K_t is a conversion

$$T = K_t \Phi I_a \tag{1}$$

and machine constant, Φ is net flux per pole, and I_a is the total armature current.

The voltage E, induced as a counter electromotive force (emf) by generator action in the parallel paths of the armature, plus the voltage drop I_aR_a through the armature due to armature current I_a and armature resistance R_a, must be overcome by the total impressed voltage V from the line. Voltage relations can be expressed by Eq. (2).

$$V = E + I_aR_a \tag{2}$$

The counter emf and motor speed n are related by Eq. (3), where K is a conversion and machine constant.

$$n = \frac{E}{K\Phi} \tag{3}$$

Mechanical power output can be expressed by Eq. (4), where n is the motor speed in rpm, T is the

$$\text{Horsepower} = \frac{2\pi nT}{33,000} = \frac{2\pi nT'}{44,742} \tag{4}$$

torque developed in pound-feet, and T' is the torque developed in newton-meters.

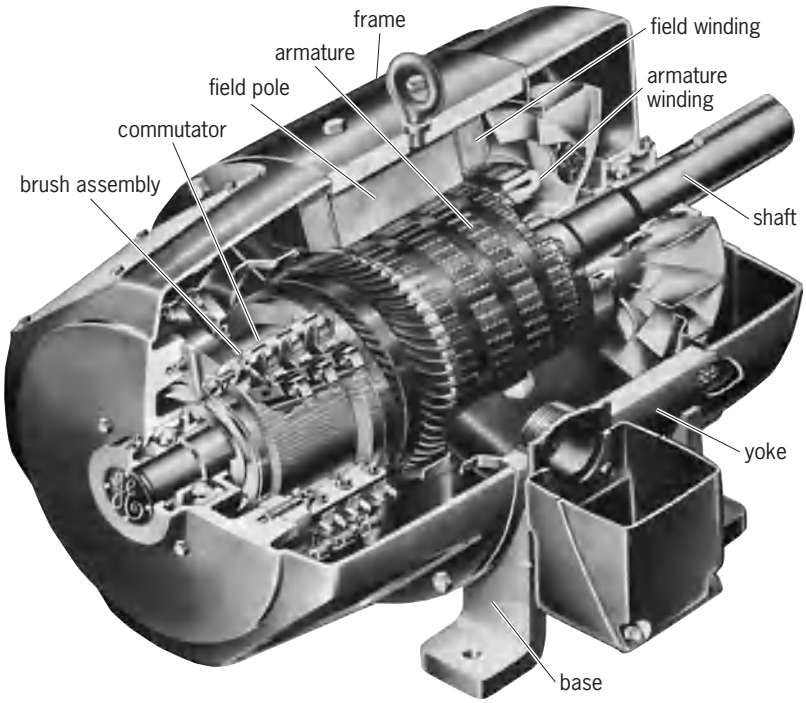

Fig. 1. Cutaway view of typical dc motor. (*General Electric Co.*)

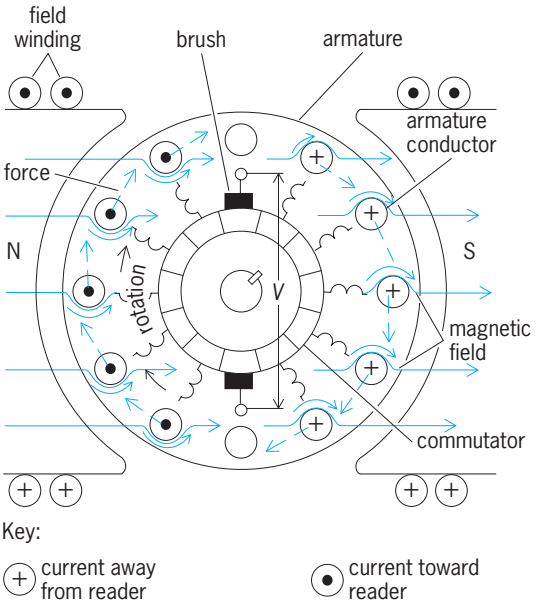

Fig. 2. Rotation in a dc motor.

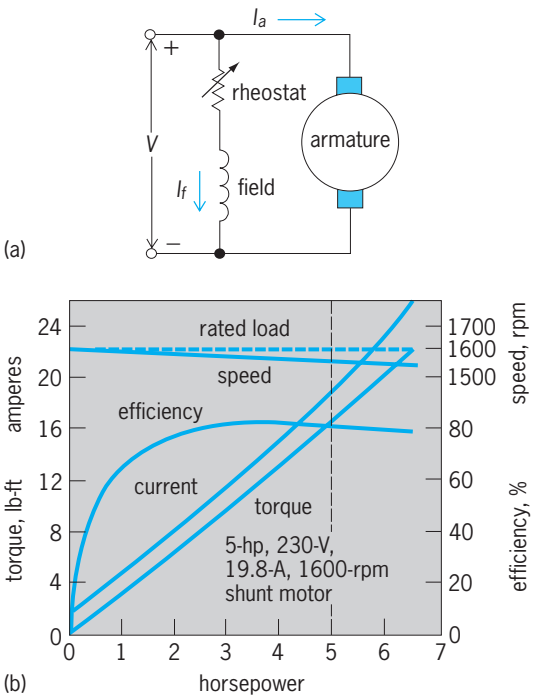

Fig. 3. Shunt motor. (*a*) Connections. (*b*) Typical operating characteristics.

By use of these four equations, the steady-state operation of the dc motor may be determined.

Types. Direct-current motors may be categorized as shunt, series, compound, or separately excited.

Shunt motor. The field circuit and the armature circuit of a dc shunt motor are connected in parallel (**Fig. 3***a*). The field windings consist of many turns of fine wire. The entire field resistance, including a series-connected field rheostat, is relatively large. The field current and pole flux are essentially constant and independent of the armature requirements. The torque is therefore essentially proportional to the armature current.

In operation an increased motor torque will be produced by a nearly equal increase in armature current, Eq. (1), since K_t and Φ are constant. Increased I_a produces an increase in the small voltage I_aR_a, Eq. (2). Since V is constant, E must decrease by the same small amount resulting in a small decrease in speed n, Eq. (3). The speed-load curve is practically flat, resulting in the term "constant speed" for the shunt motor. Typical characteristics of the motor are shown in Fig. 3*b*.

Typical applications are for load conditions of fairly constant speed, such as machine tools, blowers, centrifugal pumps, fans, conveyors, wood- and metalworking machines, steel, paper, and cement mills, and coal or coke plant drives.

Series motor. The field circuit and the armature circuit of a dc series motor are connected in series (**Fig. 4***a*). The field winding has relatively few turns per pole. The wire must be large enough to carry the armature current. The flux Φ of a series motor is nearly proportional to the armature current I_a which produces it. Therefore, the torque, Eq. (1), of a series motor is proportional to the square of the armature current, neglecting the effects of core saturation and armature reaction. An increase in torque may be produced by a relatively small increase in armature current.

In operation the increased armature current, which produces increased torque, also produces increased flux. Therefore, speed must decrease to produce the required counter emf to satisfy Eqs. (1) and (3). This produces a variable speed characteristic. At light loads the flux is weak because of the small value of armature current, and the speed may be excessive. For this reason series motors are generally connected permanently to their loads through gearing.

The characteristics of the series motor are shown in Fig. 4*b*. Typical applications of the series motor are to loads requiring high starting torques and variable speeds, for example, cranes, hoists, gates, bridges, car dumpers, traction drives, and automobile starters.

Compound motor. A compound motor has two separate field windings. One, generally the predominant field, is connected in parallel with the armature circuit; the other is connected in series with the armature circuit (**Fig. 5**).

The field windings may be connected in long or short shunt without radically changing the operation of the motor. They may also be cumulative or differential in compounding action. With both field windings, this motor combines the effects of the shunt and series types to an extent dependent upon the degree of compounding. Applications of this motor are to loads requiring high starting torques and somewhat variable speeds, such as pulsating loads, shears, bending rolls, plunger pumps, conveyors, elevators, and crushers. *See* DIRECT-CURRENT GENERATOR.

Separately excited motor. The field winding of this motor is energized from a source different from that of the armature winding. The field winding may be of either the shunt or series type, and adjustment of the applied voltage sources produces a wide range of speed and torque characteristics. Small dc motors may have permanent-magnet fields with armature excitation only. Such motors are used with fans,

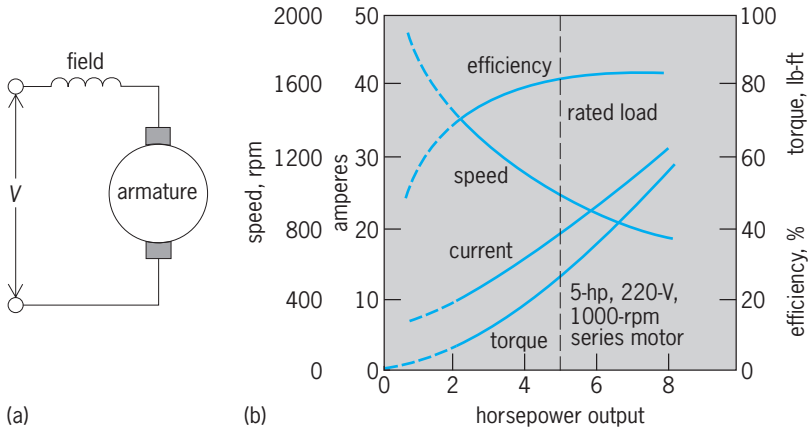

Fig. 4. Series motor. (*a*) Connection. (*b*) Typical operating characteristics.

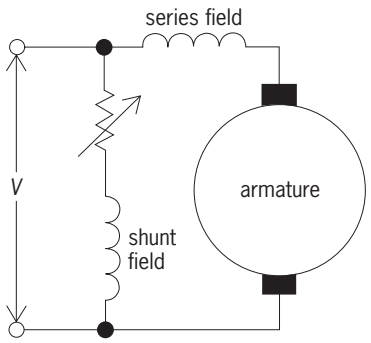

Fig. 5. Connection of a compound motor.

blowers, rapid-transfer switches, electromechanical activators, and programming devices. *See* ELECTRIC ROTATING MACHINERY. L. F. Cleveland

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Direct-current transmission

The conveyance of electric power by conductors carrying unidirectional currents. *See* DIRECT CURRENT.

Alternating current (ac) is universally accepted for the generation of electric power and for its distribution to consumers. In order to deliver power from generating stations to load centers and to facilitate interchange of power between different power companies voltages are stepped up to transmission levels. Transmission lines typically have one or two sets of three-phase ac conductors suspended from towers. Underground or submarine cables are used less frequently, in cases where the additional cost is considered justifiable for technical or environmental reasons. They are terminated at substations which house transformers to adjust the voltage and circuit breakers to establish appropriate connections. *See* ALTERNATING CURRENT; ELECTRIC POWER TRANSMISSION; TRANSMISSION LINES.

Power companies face increasing pressure to reconcile increasing load demands with financial, environmental, and resource constraints. This emphasizes the need to operate economically and effectively while preserving reliability. Consequently, it is often preferable to interconnect systems into larger power pools, even crossing national boundaries, rather than seek total self-reliance on exclusive generation capability within each company. *See* ELECTRIC POWER SYSTEMS.

Within this established pattern of ac power systems, application of transmission equipment operating with direct current (dc) has increased since the late 1950s. Although the adoption of dc over a century ago by Thomas Edison for an entire system from generation to load has long been superseded by ac practice, special circumstances arise where dc transmission has economic or technical advantages. Thus dc becomes a medium for interconnecting points in ac systems. Special converters within converter stations change the ac to dc (rectification) and dc back to ac (inversion) to complete the power interchange.

Rationale. A dc line with two conductors is cheaper to construct and often has lower power losses than a three-phase ac line rated for the same power. Moreover, the same dc line is often considered as equal in reliability of service to a double-circuit three-phase line. The economic advantages are proportional to the line length but are offset by the substantial cost of the converting equipment. On this criterion alone, there would be a break-even distance beyond which dc would be cheaper: several hundred kilometers for overhead lines and much less for more costly underground and submarine cables. Indeed, this has been an important factor in a small minority of projects. The break-even distance is specific to the conditions and financing of each project, and it obviously contradicts the zero distance of back-to-back installations. However, several other factors influence the selection of dc.

If the ac frequencies at the converting stations are nominally the same but controlled separately, their frequency independence is maintained by the dc link. In other words, the dc system is an asynchronous link. This is the justification of many back-to-back schemes such as the ties between regions in the United States and between European countries.

Although North America operates at 60 Hz and most other parts of the world operate at 50 Hz, on occasion there is a need to interconnect ac systems having different nominal frequencies. The asynchronous nature of dc serves as a frequency changer with the control action of each converter synchronized to its local ac frequency. The Sakuma frequency changer in Japan is a back-to-back scheme, connecting two regions which, for historical reasons, have frequency standards of 50 and 60 Hz.

The electrical shunt capacitance of cables is charged and discharged at the frequency of the voltage. Since the capacitance is proportional to distance, an ac cable longer than a few tens of miles is loaded to its thermal rating by a capacitor charging current, with no power being conveyed to the remote termination. Unless the cable can be sectioned for intermediate compensating measures, dc is obligatory for many cable applications. This is especially the case for submarine links where overhead lines are not an option and economic considerations of break-even distance may or may not have favored dc.

All parts of an ac system function at the same nominal frequency. A system is designed to ensure that the generators do not lose synchronism despite load variations and large fault disturbances. The system is then considered to be dynamically stable. This becomes more of a challenge when the generating stations are geographically dispersed and remote from load centers, as opposed to the relatively tightly knit

systems found in Europe. Should an ac connection to another system be contemplated, the combined system is intended to operate in synchronism, although the dynamic stability in either system may deteriorate below acceptable security. In comparison, a dc interconnection maintains the dynamic independence of each system. For example, remote hydroelectric generation at Churchill Falls in Labrador and on rivers entering James Bay prevents Hydro-Quebec from establishing synchronous connections with neighboring power companies for reasons of potential instability. Mutual interties and export contracts for electric power to the New York power pool and New England utilities have been implemented by several dc links. The question then turns to the selection of power rating and the location of the dc terminal stations.

The automatic control circuits at the converting stations permit the dc power to be accurately set at a value determined by the system control center. Furthermore, the dc power is maintained during dynamic ac frequency disturbances and can be rapidly changed (modulated) in as little as a few milliseconds, on demand. The same control permits the direction of dc power to be reversed equally quickly. This ability to ride through disturbances and to permit precise power scheduling and modulation responsive to dynamic needs has become of increasing value in the operation of power systems. The power flow in an individual ac line cannot be independently controlled to the degree offered by dc transmission.

Applications. In most instances, manufacturers bid internationally to supply equipment adapted to the needs and specification of each project. If the project is not contingent on building generating stations, a pressing need may reduce the interval between conception and operation to as little as 2 years, although, as with many large power projects, the process can take several years and the transmission capacity may be scheduled in stages. Systems in operation around the world are of varying lengths, operating voltages, transmission capacities (megawatts), and applications. There are 77 systems in service and another 16 either under construction or under active consideration. Of the total, 25 are in the United States and Canada; the remaining 68 involve 27 other countries. Compared with the first system established in 1954, which supplies the island of Gotland with 30 MW of power from the Swedish mainland, the total nominal power rating of dc transmission from the Nelson River hydroelectric generation in Manitoba, Canada, is 3420 MW.

Categories of application are (1) long-distance overhead dc lines conveying bulk power from remote generating stations to load centers, such as Itaipu, Brazil; (2) long-distance seasonal or daily power interchange, such as the Pacific DC Intertie linking Oregon with southern California; (3) submarine cables, such as for Cross Channel 2 between Great Britain and France; (4) zero-distance connections (back-to-back schemes) between adjacent systems, such as Miles City, Montana; and (5) underground dc cables in urban areas, such as in Kingsnorth, London.

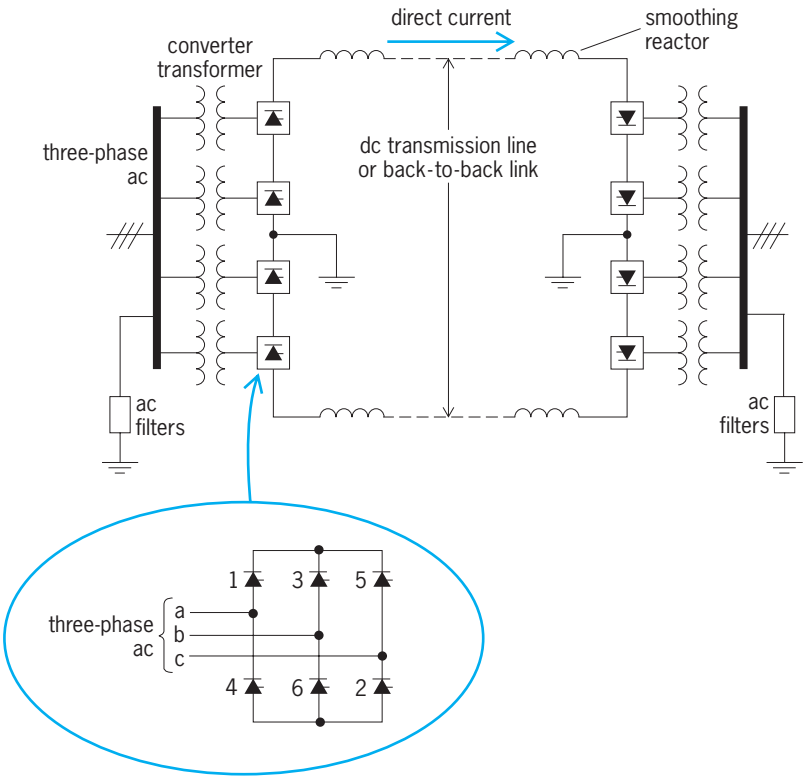

Fig. 1. Diagram of typical dc system.

Equipment. A typical dc converter station contains conventional ac equipment in its ac switchyard supplemented by equipment specific to the ac-dc conversion. Solid-state converters are connected on the dc side with a center neutral point which is usually connected to a remote ground electrode. Balanced dc currents are circulated on each pole at plus and minus dc voltages with respect to ground (**Fig. 1**). *See* ELECTRIC POWER SUBSTATION.

Each converter is connected to the ac system through transformers which, in conjunction with the control circuits, determine the dc pole voltages. Filters remove ac harmonics arising from the conversion process, and the dc current is smoothed by a reactor which may be supplemented by more dc filter components. Filtering attenuates potential radio and telecommunication interference as well as avoiding deterioration of the ac waveforms. *See* ELECTRIC FILTER.

Although mercury-arc converters were applied until 1974, it is now the practice to use solid-state devices. The converters are made up of groups of Graetz-connected bridges, each containing six valves. Each valve uses a matrix of thyristors which are simultaneously gated into conduction through fiber optics in the firing sequence indicated by the numbers in the inset of the illustration. Each thyristor can withstand several kilovolts in its nonconducting state and several thousand amperes when conducting. Although the current rating is often sufficient without connecting thyristors in parallel, many must be connected in series to achieve the voltage

withstand requirements. The result is a unidirectional valve which is controlled to turn on once per ac cycle and which turns off when its voltage reverses. Depending on the points in the cycle at which the valves are sequentially fired, the dc voltage is controlled from rectification through to inversion when the power flow is reversed and the dc converted back to ac. The six-valve bridges are often connected as series pairs in 12-pulse groups with a staggered firing sequence in order to produce a smoother dc waveform and reduced ac harmonics. *See* COMMUTATION; CONVERTER; SEMICONDUCTOR RECTIFIER.

Generally, the converters are indoors, with the connections to outdoor converter transformers made via insulated conductors (bushings) through the walls of the valve hall. The bushings are often integral parts of the transformer structures. *See* TRANSFORMER.

Except in back-to-back schemes, the converters are connected through the dc switchyard to the lines interconnecting the converter stations. For overhead lines, conductors for the two poles of opposite polarity of voltage are typically carried on the same tower. Balanced pole currents avoid ground current. In principle, ground can serve as a return-current conductor. In practice, local restrictions on the size and duration of ground current due to concerns for effects on buried pipelines often limit monopolar operation to short emergency periods. In order to achieve an effective ground connection capable of carrying, even temporarily, large currents, a separate insulated overhead line or cable is run from the converting station to a suitable ground-electrode site. The possibility of rapidly discontinuing dc current by converter control avoids the need for dc circuit breakers to deenergize dc line faults.

Trends. As a highly innovative technology, dc transmission has broadened its scope of application along with parallel development of equipment. Valves apply state-of-the art power semiconductors with advanced cooling and gating methods. Control circuits operate in real time and are now computer-based. They are highly responsive to various control strategies within the systemwide control, extending back to the system control center. The dc substations are typically interlinked by microwave communication. Circuit breakers are available to interrupt large dc currents at high voltages encountered in dc systems and can be used for special purposes. The ratings of individual thyristors have increased, resulting in lower converter sizes and costs. Large power thyristors can be triggered by a light pulse directly into the thyrstor, resulting in simplification of the electronic gate firing circuits. *See* CIRCUIT BREAKER.

A significant advance is the extension of point-to-point two-terminal practice to the construction of dc networks containing three or more converter stations for a coordinated power interchange in a so-called multiterminal system. It has been achieved by tapping the existing Italy-Sardinia connection to feed power to Corsica. A five-terminal network takes power by dc from James Bay through Quebec into New England.

China, India, and developing countries in southern Africa are actively exploring the possibility of increasing the dc voltage rating from the present maximum of ±600 kV (the Itaipu project in Brazil) to ±800 kV. This would allow power flow on one transmission line to approach 6000 MW, enough power to supply a population of a few million. However, building a transmission line of such large capacity faces challenges, including maintenance of service if the line should be unexpectedly removed from service due to adverse environmental conditions.

The blackouts in the northeastern United States and Canada (August 15, 2003) and Italy (September 28, 2003) have reawakened the idea that a blackout could be blocked from spreading by dividing the power system into segments and inserting back-to-back dc transmission units in the ac power lines wherever they cross-segment boundaries. Quebec interfaces to the northeastern U.S. power system through dc transmission, which prevented the blackout of August 15, 2003 from affecting it.

As the price of fossil fuels increases, causing electric power prices to rise as well, there are incentives to search for renewable energy. Hydroelectricity, wind, solar, and geothermal energy as well as less expensive energy sources such as some forms of coal will spur development of long-distance electric power transmission between energy sources and load centers. It is expected that high-voltage dc transmission will play a vital role in helping deliver less expensive energy to where it is most needed. *See* ELECTRIC POWER GENERATION; ENERGY SOURCES.

VSC transmission. Advanced high-voltage dc transmission facilities are being built using voltage sourced transmission or voltage source converter (VSC) transmission (**Fig. 2**). These new developments apply the VSC converter bridge, which uses gate turnoff thyristors (GTOs) and insulated gate bipole transistors (IGBTs). Its special properties include the ability to independently control real and reactive power at the connection bus to the ac system. Reactive power can be either capacitive or inductive and can be controlled to quickly change from one to the other.

A voltage source converter that functions as an inverter does not require an active ac voltage source to commutate into as does the conventional line commutated converter. The VSC inverter can generate an ac three-phase voltage and supply electricity to a load as the only source of power. It does require harmonic filtering, harmonic cancellation, or

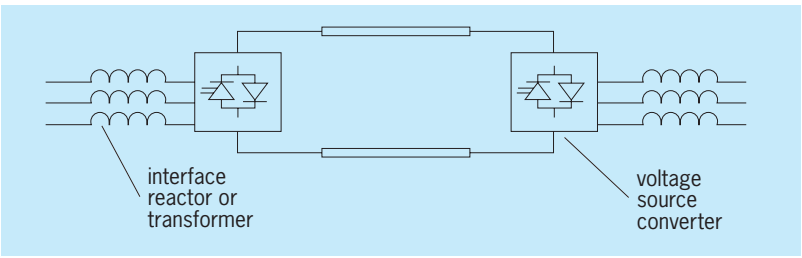

Fig. 2. Simplified diagram of VSC transmission.

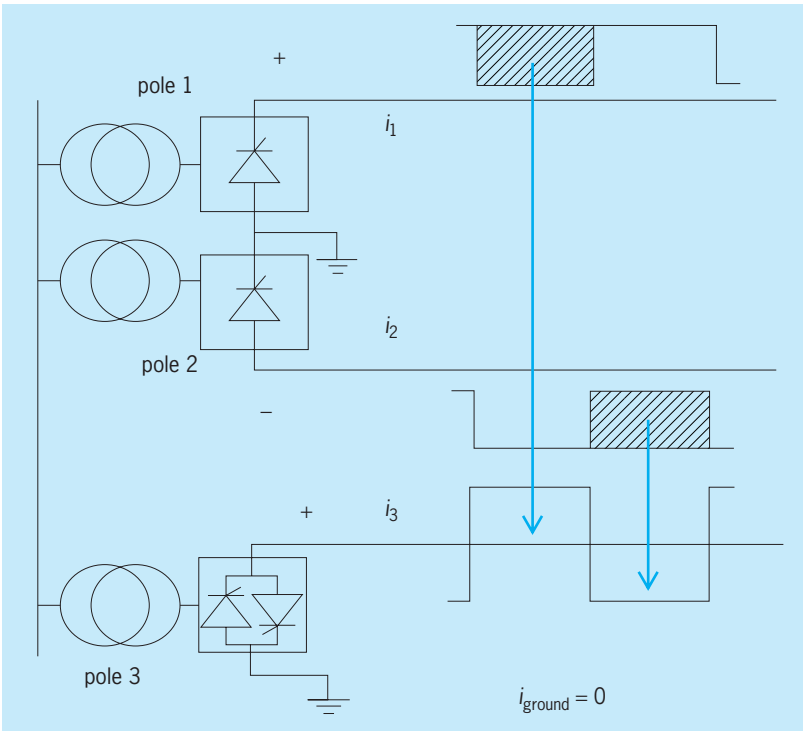

Fig. 3. Configuration of a tripole showing how current flows on each pole conductor.

pulse-width modulation to provide an acceptable ac voltage waveshape.

The first VSC transmission system went into operation in 1997 in Sweden. VSC converter bridges can be applied in back-to-back configuration and for major transmission systems. The back-to-back VSC link is the ultimate transmission and power flow controller. It can control and reverse power flow easily, and control reactive power independently on each side. With a suitable control system, it can control power to enhance and preserve ac system synchronism, and act as a rapid phase-angle power flow regulator with 360° range of control. The first North American back-to-back VSC unit was the Eagle Pass 36-MW interconnection between Mexico and Texas commissioned in 2000. For point-to-point transmission systems, power levels reaching 1100 MW and greater are possible. A 180-MW interconnection using underground cable between Queensland and New South Wales in Australia was completed in 2000. Other installations include a 330-MW under water interconnection between Long Island and Connecticut.

There is considerable flexibility in the configuration of VSC converter bridges. Many two-level converter bridges can be assembled with appropriate harmonic cancellation properties to generate acceptable ac system voltage waveshapes. Another option is to use multilevel converter bridges to provide harmonic cancellation. Both two-level and multilevel converter bridges can utilize pulse-width modulation to eliminate low-order harmonics. However, with pulse-width modulation, high-pass filters may still be required since higher-order harmonics are then generated.

The great benefit derived from inexpensive power transmission cable is environmental, although relative ease of licensing is also an advantage, since licensing of less expensive, overhead transmission lines is difficult and time consuming. Cables for VSC transmission are cheaper to bury than conventional power cables, and the lead time for licensing can be significantly reduced, allowing the transmission to be brought into service sooner than is possible for overhead transmission lines. For some locations it may be impossible to build an overhead transmission line, but buried cable may be an acceptable, although relatively expensive, option.

The maximum ratings for VSC transmission in one cable transmission circuit so far are ±300 kV and 1850 amps. This should increase in the future.

Bidirectional valves. Current flows through conventional high-voltage dc thyristor converter valves in one direction only. It is possible to construct a valve with either bidirectional thyristors, or two thyristor strings in parallel but with opposite current-conducting directions. Under controlled operation, converter bridges with bidirectional valves provide additional benefits, the most significant being three-pole (tripole) operation without a neutral return path. Such operation can be applied to achieve the maximum power in the three phases of an ac line if it is converted to dc. Tripole dc transmission is a new concept that is being actively studied.

The unique operation of tripole is based on having two poles in bipole configuration (**Fig. 3**). The positive pole raises its current to an overload value for several minutes while the negative pole current is less than rated current. The current in the third pole makes up the difference with its voltage polarity maintaining the direction of desired power flow. After several minutes, the current in the first pole reduces to allow its conductor to cool, and the current in the second pole picks up to the overload value at which the first pole was operating. The transition of current between overload and light load levels in the first and second poles requires that the current in the third pole reverse both its current direction and voltage polarity in order that power flow in the three poles remains steady. Hence, the valves in the third pole must be bidirectional.

The converters have no overload capacity above their rating and so the valves of the first and second poles must be rated to the overload level. The transmission or cable conductors, on the other hand, can easily withstand brief current overloads. If the period of overload and light load in these poles is of the order of several minutes, conductor temperature variations are minimal, of the order of several degrees, allowing the cyclic overload operation to continue. The current level in the third pole does not need to overload, but it does change polarity. The transition period is over a short period of time of several seconds under controlled conditions that maintain steady power transmission.

Significantly more power can be made to flow on a long ac transmission line when it is converted to dc in this way than is possible if the transmission

line remains ac. Consequently, maximum use can be made of existing transmission lines.

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Direction-finding equipment

Equipment used for determining the direction from which a received signal is being radiated. Direction finders serve a variety of applications. Direction finders can be used for navigation purposes to determine the heading of a vehicle relative to a transmitter at a known location, and to determine the location of the vehicle when the range to that transmitter is known, as in the case of VOR/DME instrumentation. Direction finding may also be used to determine vehicle heading for guidance along a desired route toward a given destination. Surveillance uses of direction finders include discovering and locating covert transmitters used for espionage, locating the emitter in search-and-rescue missions, and tracking crewless objects in scientific missions. A direction finder may be implemented either as a ground station used to track a moving source or on a moving vehicle. Bearing information obtained by the ground station can also be passed to a moving vehicle that is equipped with a transmitter that can be tracked by the ground station. As mentioned above, a direction finder may also be found on a moving vehicle, such as an aircraft or a ship, to determine the vehicle's heading relative to a transmitter or the transmitter's bearing relative to the known track of the vehicle. *See* DISTANCE-MEASURING EQUIPMENT; VOR (VHF OMNIDIRECTIONAL RANGE).

Principles of direction finding. Most direction-finding systems are based on one or the other of two principles, where one exploits antenna gain characteristics and the other exploits phase and frequency relationships observed by a moving or a multiple-element fixed antenna.

Antenna gain. Direction-finding systems based on antenna gain take advantage of variations of the antenna gain pattern, whereby the received signal amplitude varies from a maximum to a minimum according to its angle of arrival. Signal-minimum systems exploit the distinct null found in an antenna's gain pattern. Some of the earliest direction-finding equipment using loop antennas was conceived on this principle. The advent of higher-frequency systems and highly directive antennas, such as phased arrays and parabolic dishes, provided the basis for signal-maximum systems, which can pinpoint the direction in which the signal is strongest. *See* ANTENNA (ELECTROMAGNETISM).

Induced Doppler frequency modulation. The principle of induced Doppler frequency modulation requires movement of the antenna to cause additional Doppler shift to the original signal. The antenna motion is generated in such a way that the resulting shift in received frequency will provide information about the direction of the incoming signal with respect to that of the known antenna motion. The frequency of the signal received by the moving antenna will vary in relation to that received by a fixed antenna, the latter serving as a reference signal to extract the content induced by the moving antenna. The antenna motion most suited for omnidirectional purposes is circular. For this, the induced Doppler is zero at the instant when the antenna motion is perpendicular to the bearing of the received signal. *See* DOPPLER EFFECT.

Airborne automatic direction finder. Airborne automatic direction finders (ADFs) are constrained by size and weight. For their simplicity, signal-minimum systems with loop antennas are widely used. Although there are antenna configurations that use a continuously rotating or steerable antenna to optimize the reception of a signal from any direction, a direction finder with orthogonal fixed-loop antennas has the advantage of not having any mechanically moving parts in the antenna assembly, reducing maintenance requirements and improving reliability. The signals from the loop antennas carry bearing information in their relative amplitudes as determined by the loop-antenna gain pattern (**Fig. 1**).

The patterns in Fig. 1 illustrate the antenna response to a plane-wave signal that arrives at a bearing of β relative to vertical. The loop antenna cannot determine whether the signal arrives at β or at $\beta + 180^\circ$. Therefore, an additional (omnidirectional) antenna is usually provided to fully resolve the bearing measurement. To extract this information, the signals from the loop antennas are amplitude modulated separately with modulation signals that are mutually orthogonal in phase. When the amplitude-modulated signals are combined, the resulting signal has a specific phase relationship with respect to the original modulation signal. This relationship is proportional to the bearing β of the received signal, with an ambiguity of 180°. However, the signal

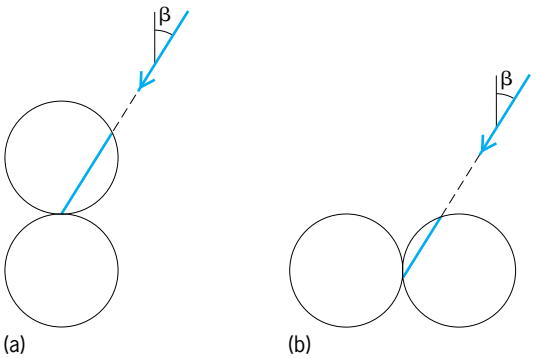

Fig. 1. Signal amplitude differences between orthogonal loop-antenna patterns. (*a*) Signal amplitude = *A* cos β. (*b*) Signal amplitude = *A* sin β.

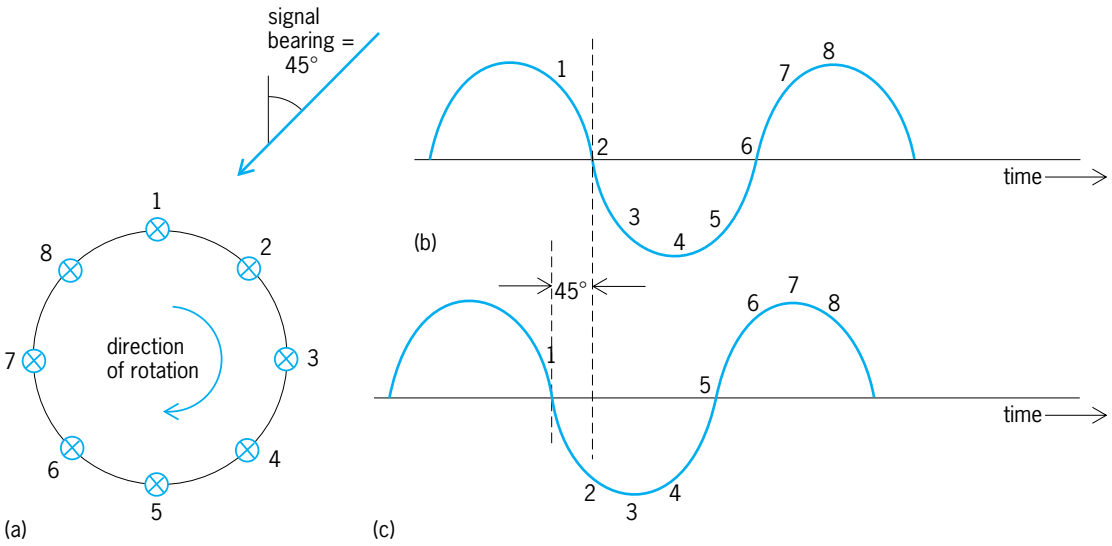

Fig. 2. Induced Doppler direction finder using circular antenna motion. (a) Diagram of equipment with sequential switching of antennas 1–8. (*b*) Extracted Doppler profile, which goes to zero at locations 2 and 6, where the antenna is moving orthogonally to an incoming signal with bearing of 45°. (*c*) Reference profile based on antenna motion.

from the omnidirectional antenna is also added to the amplitude-modulated signals from the loop antennas, contributing what is known as the sense term to the combined signal.

After synchronous demodulation to move the time dependence of the received signal, a phase comparison of the baseband signal with the original modulation signal will yield the value of β. Without the sense term, the arbitrary value of the residual phase error would produce a 180° phase ambiguity to the solution of β. *See* AMPLITUDE-MODULATION DETECTOR.

Doppler direction finder. A Doppler direction finder using circular antenna motion induces a sinusoidal frequency variation to the original frequency content of the received signal (**Fig. 2***a*). To obtain the bearing information, the signal first needs to be demodulated with a frequency-modulation (FM) discriminator. The discriminator output is a sinusoid (Fig. 2*b*) whose phase is shifted with respect to the reference signal (Fig. 2*c*) by an amount that corresponds to the bearing of the received signal. A post-FM demodulator is needed to extract that phase angle.

In most implementations of this type of direction finder, the antenna motion is synthetically generated by sequentially switching through a series of antennas arranged in a circle (Fig. 2*a*). The measurements made from each antenna provide information that can be implicitly reconstructed according to principles of signal sampling.

Accuracy factors. Direction finders based on antenna gain depend on the fidelity of the gain pattern. Pattern imperfections will affect the accuracy of the derived bearing. All direction finders are susceptible to contamination of the true signal with a spurious reflection, causing measurement error. Low- and medium-frequency direction finders are particularly prone to sky-wave contamination. *See* AIR NAVIGATION; ELECTRONIC NAVIGATION SYSTEMS.

Patrick Hwang; Richard L. Greenspan

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Directional coupler

A four-port waveguide device (**Fig. 1**) in which an incoming wave at any one port (for example, A) appears at two others (C and D) but not at the fourth (B). This device finds numerous applications in waveguide networks such as microwave waveguides, integrated optics, and optical fibers. It is the equivalent of the hybrid induction coil used in conventional wire circuit telephony to provide side-tone balance. *See* TELEPHONE.

In Fig. 1, P_A is the input power, and P_B, P_C, and P_D are the powers of the output waves at ports B, C, and D respectively. Ideally, $P_B = 0$, and there is

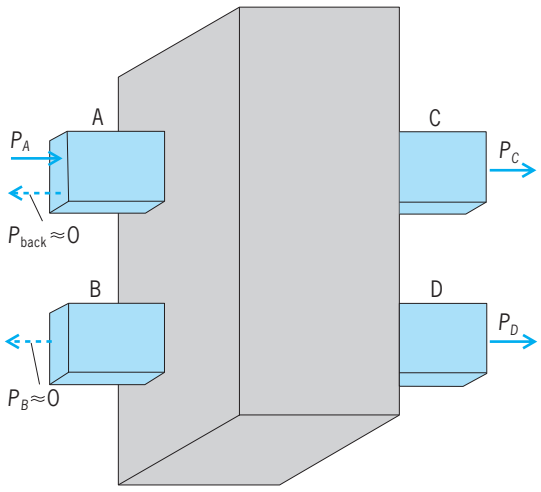

Fig. 1. Four-port directional waveguide coupler.

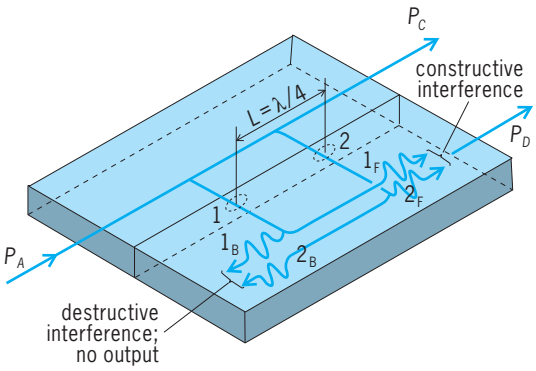

Fig. 2. Two-hole directional coupler.

no spurious reflection from the input port: $P_{back} = 0$. As a system component, the performance of a directional coupler is quoted in terms of (1) its coupling factor, 10 $\log_{10}$ (P_D/P_A), in decibels (dB); (2) directivity, 10 $\log_{10}$ (P_D/P_B), in dB; (3) insertion loss, 10 $\log_{10}$ $[(P_C + P_D)/P_A]$; (4) frequency band of operation; and (5) its matching impedance as measured by any spurious backward power flow from input port A. *See* ATTENUATION (ELECTRICITY); DECIBEL; ELECTRICAL IMPEDANCE; IMPEDANCE MATCHING.

Two-hole coupler. Directional couplers are usually made by introducing some form of lumped or continuous coupling among several guides so that the coupled waves interfere constructively in some directions while canceling one another in other directions. Consider two identical microwave waveguides running adjacent to each other with two small holes of the same size and shape coupling between them (**Fig. 2**). These two holes will divert a small portion of the incoming wave at port A from the left-hand waveguide into the right-hand waveguide, thereby establishing four propagating components. The two forward-propagating components 1_F and 2_F have traveled different paths but equal distances from port A to port D. They are, therefore, always in phase and will combine constructively to produce the output wave at port D. As for the backward-propagating components, 1_B leads 2_B by a phase of $4\pi L/\lambda$ radians, where L is the distance between the two holes and λ is the wavelength of the propagating radiation. Evidently, if $L = \lambda/4$, 1_B and 2_B are 180° out of phase, and since they have the same amplitudes (due to the small-hole assumption), they will cancel each other and send no power to port B. *See* INTERFERENCE OF WAVES.

Multipath and slot couplers. The relation $L = \lambda/4$ and the smallness of the holes impose two limitations on the performance of the simple two-hole directional coupler: a narrow-frequency band over which the directivity is high, and a small coupling factor which is on the order of 10–30 dB; that is, $(P_D/P_A) = 10^{-1}$ to 10^{-3}. To alleviate these problems, multipath couplers are used, where the backward cancellation is obtained through destructive interference among waves excited by three or more coupling elements (such as holes), usually spaced a quarter-wavelength apart. The coupling holes are not necessarily equal. Their coupling is frequently arranged to be in a binomial series (for example, five holes with couplings proportional to 1:4:6:4:1) or according to some other tapering law. A natural extension to the multipath coupler is the slot coupler, where the coupling element is a continuous slot that is several wavelengths long and often tapered, and the coupling is continuously distributed along the length of the interaction. *See* BINOMIAL THEOREM.

Couplers for open waveguides. Holes and slots are appropriate coupling elements for closed metallic waveguides, where the propagating field is completely enclosed within the walls of the guide. Other design alternatives exist for open waveguides such as microwave striplines, diffused dielectric optical guides used in integrated optics, and optical fibers. In these guides, the propagating field extends beyond the physical dimensions of the guiding structure (for example, the metallic plates of the stripline or the core of the fiber), and proximity coupling becomes possible.

Figure 3 shows a possible arrangement for a single-mode fiber-optic directional coupler. In a single- mode cylindrical fiber, optical power is carried in a very thin (5–10-micrometer diameter) core whose index of refraction is slightly higher than that of the surrounding cladding material. Similar to other dielectric waveguides, the electromagnetic field is not entirely confined within the core, and an exponentially decreasing, evanescent, nonradiating tail exists in the first few micrometers of the cladding. Fusion, etching, and mechanical polishing techniques may be used to bring two fiber cores close enough to each other to make possible the transfer of power from one core to the other through the interaction of the corresponding evanescent fields. The long interaction length (with respect to an optical wavelength) ensures an extremely high directivity (greater than 60 dB), and the gradual approach of the cores to each other virtually eliminates back reflections. As the separation between the cores decreases, a larger and larger percentage of the input power is coupled to port D, and a complete power transfer ($P_D = P_A$ and $P_C = 0$) is easily achieved. For even stronger coupling (smaller separation or longer interaction length), the power which was coupled to

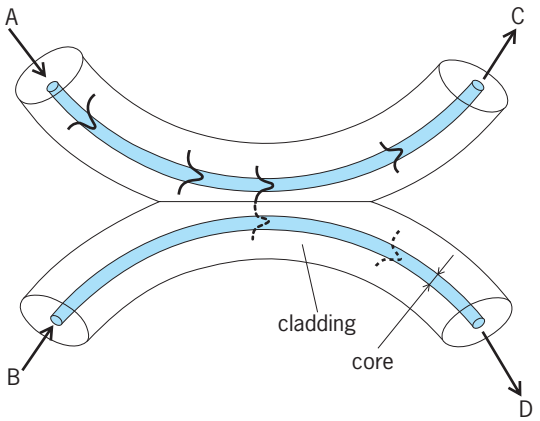

Fig. 3. Single-mode fiber-optic directional coupler.

the lower fiber may couple back to the upper fiber, a condition known as overcoupling.

Relation of output to input. Assume now that input waves are present at both port A and port B. Since directional couplers are linear devices, the output fields are linear combinations of the input fields. For simplicity, only the scalar case is considered where the input as well as the output waves have the same state of polarization. Thus, if E_A, E_B and E_C, E_D are, respectively, the amplitudes of the input and output fields, they are related to each other through Eqs. (1), or, in a matrix form, through Eq. (2). (The $\{M_i\}$ are, in general,complex numbers.)

$$\begin{aligned} E_C &= M_1 \cdot E_A + M_2 \cdot E_B \\ E_D &= M_3 \cdot E_A + M_4 \cdot E_B \end{aligned} \tag{1}$$

$$\begin{bmatrix} E_C \\ E_D \end{bmatrix} = \begin{bmatrix} M_1 & M_2 \\ M_3 & M_4 \end{bmatrix} \cdot \begin{bmatrix} E_A \\ E_B \end{bmatrix} \tag{2}$$

For a lossless coupler, the matrix M is unitary; that is, the inverse of M is also the complex conjugate of M. When the coupler is also symmetric, as in Figs. 2 and 3, M has a much simpler form, given by Eq. (3),

$$\begin{bmatrix} E_C \\ E_D \end{bmatrix} = \begin{bmatrix} \sqrt{1-K} & j\sqrt{K} \\ j\sqrt{K} & \sqrt{1-K} \end{bmatrix} \cdot \begin{bmatrix} E_A \\ E_B \end{bmatrix} \tag{3}$$

where K is the linear coupling factor and $j = \sqrt{-1}$. There is a phase difference of of 90° between the coupled portion of the input wave (M_3, E_A) and the uncoupled component (M_1, E_A). This property, which is also shared by bulk optic beam splitters, the waveguide magic tee, and the hybrid induction coil, is crucial for the understanding of the energy balance in various interferometers. *See* INTERFEROMETRY; MATRIX THEORY.

Applications. Directional couplers, whether fixed or variable, are extensively used in microwave and optical circuits. Weak couplers can selectively monitor either the forward or backward power flow in waveguides, and medium-to-strong couplers serve as amplitude combiners and splitters in various resonators and sensors. All of these applications require high directivity, low insertion loss, and negligible back reflections. Many other useful junctions are also based on the same physical principles. *See* INTEGRATED OPTICS; MICROWAVE MEASUREMENTS; MICROWAVE POWER MEASUREMENT; OPTICAL COMMUNICATIONS; OPTICAL FIBERS; WAVEGUIDE. Moshe Tur

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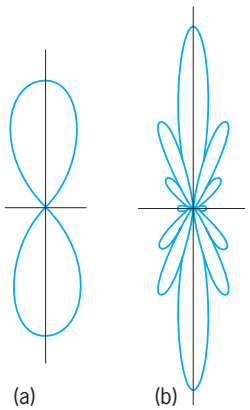
Directivity

The general property of directional discrimination displayed by systems that receive or emit waves. Thus, loudspeakers, microphones, radio antennas, underwater sound projectors, hydrophones, and even telescopes all have the common property that their effectiveness depends upon the direction from which the wave is either emitted or received. The manner in which a sender or receiver is directional depends largely upon its geometrical shape and, in particular, upon its dimensions compared to those of the wavelength involved (λ).

Directivity is a desirable property of a receiver because it permits the identification of the direction from which a signal comes and because noise from other directions is eliminated. It is desirable in a sender because the available energy can be concentrated in a given direction. A simple example of directivity is furnished by the megaphone, which effectively increases the size of the emitting area, thus increasing directivity.

Directivity is specified in terms of plane waves of a given wavelength in the presence of isotropic noise. The directional property of a projecting system is described quantitatively by a directivity factor D, which is the ratio of the intensity of the signal at a point on a designated axis to that at the same point produced by a nondirectional source radiating the same total power. The directivity index DI, given in decibels, is 10 log 10D. The directivity of some sources can be calculated, but complicated ones must be measured. Most transducers have the same directivity factor whether they are used as projectors or receivers. *See* DECIBEL.

In the case of receiver arrays, directivity arises because the various elements respond to the incoming wave in different phase. As an example, consider a receiver consisting of two elements, each small compared to λ (and therefore omnidirectional by themselves), which are separated by a distance d. A plane wave incident so that the wavefront is parallel to the axis joining the receivers has the same phase at both elements, regardless of the value of λ, and the response is always a maximum. The response for a wave coming along the axis depends on the ratio of d to λ. If $d = \lambda/2$, the response is zero because the two elements receive signals which are out of phase. The polar directivity pattern for this case is shown in **illus.** *a*.



Polar directivity patterns (*a*) for a dipole with the spacing $\lambda/2$; (*b*) for linear array of seven elements spaced $\lambda/2$ apart.

If more elements are added, all spaced $\lambda/2$ apart, directivity is increased and side (or minor) lobes appear. Side lobes are directions of preferential response of lower sensitivity than that of the central direction. The multiple-element case is shown in illus. *b*. Advanced array design theory provides methods of reducing the side lobes still further by shading, or varying the amplitude and phase of the elements in a complicated way. Increased directivity can also be obtained by using acoustic lenses or parabolic reflectors. Arrays can normally be steered, either by physically rotating the array, or by introducing an appropriate phasing network for each element and steering the unit electrically.

The cases just described are samples of linear arrays in which directivity is obtained only with respect to one angular direction. Directivity in both angles (that is, where a beamlike pattern is formed) is obtained by using a receiver having extension in two dimensions. A common example of this is a plane circular piston. If the ratio of piston diameter to λ is increased, the main response is narrowed and the side lobes are reduced. For a discussion of the directivity of electromagnetic transducers *see* ANTENNA (ELECTROMAGNETISM); PARAMETRIC ARRAYS; SOUND; UNDERWATER SOUND.

William S. Cramer

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Disasteroida

A paraphyletic grouping of extinct irregular echinoids, members of which gave rise to the Holasteroida and Spatangoida in the Lower Cretaceous. Disasteroids have a split apical disc; the posterior two ocular plates and their associated ambulacra are separated from the remainder of the apical disc (and the anterior three ambulacra) by intercalated interambulacral plates. The apical disc became split in the early evolution of this group as a consequence of periproct migration out of the apical disc. Disasteroids have a strong bilateral symmetry. The peristome, which is small and oval and lacks buccal notches, lies toward the anterior of the lower surface. Plastron plating is usually protosternous but in late members can be meridosternous. Tuberculation is fine, dense, and uniform, without regional differentiation. Larger pore pairs, arranged in phyllodes around the mouth, suggest that disasteroids were deposit feeders with penicillate feeding tube feet. Most species probably lived infaunally, though not burrowing deeply.

Five families are included: Acrolusiidae, Collyritidae, Disasteridae, Pygorhytidae, and Tithoniidae, separated principally on apical disc plating. Approximately 75 species have been named, arranged in 17 genera. The oldest species is late Bajocian (Middle Jurassic) and the youngest Albian (Lower Cretaceous). The group achieved its greatest diversity during the Upper Jurassic. *See* ECHINODERMATA.

Andrew B. Smith

Discriminant

For a polynomial $f(x) = a_n x^n + a_{n-1}x^{n-1} + \cdots + a_1 x + a_0$, the discriminant is given by the expression

$$D = a_n^{2n-1}(x_1 - x_2)^2(x_1 - x_3)^2 \cdots (x_1 - x_n)^2(x_2 - x_3)^2 \cdots (x_2 - x_n)^2 \cdots (x_{n-1} - x_n)^2$$

where $x_1, x_2, \ldots, x_n$ are the roots of the equation $f(x) = 0$. There are $n(n-1)/2$ terms $(x_i - x_j)^2$ in the product corresponding to all possible selections of two indices i less than j from the numbers $1, 2, \ldots, n$.

The importance of the discriminant D lies in the fact that D vanishes if, and only if, the equation $f(x) = 0$ has equal roots. Since the value of the discriminant is unchanged if any two letters x_i and x_j are interchanged, it is a symmetric function of the roots, and can be expressed in terms of the coefficients of $f(x)$. Such an expression for D is most easily obtained, using the result $D = [(-1)^{n(n-1)/2}/a_n] \times R_x(f, f')$, where $f' = f'(x)$ is the derivative of $f(x)$ and $R_x(f, f')$ is the resultant of $f(x)$. For example, with $f(x) = a_2x^2 + a_1x + a_0$, and $f'(x) = 2a_2x + a_1$, notation (1) is obtained, and $D = a_1{}^2 - 4a_2a_0$, which

$$R_x(f, f') = \begin{vmatrix} a_2 & a_1 & a_0 \\ 2a_2 & a_1 & 0 \\ 0 & 2a_2 & a_1 \end{vmatrix} = -a_2a_1{}^2 + 4a_2{}^2a_0 \quad (1)$$

is the discriminant of a quadratic polynomial. The discriminant of a cubic polynomial $f(x) = a_3x^3 + a_2x^2 + a_1x + a_0$ is shown in notation (2).

$$D = 18a_3a_2a_1a_0 - 4a_2{}^3a_0 + a_2{}^2a_1{}^2 - 4a_3a_1{}^3 - 27a_3{}^2a_0{}^2 \quad (2)$$

See EQUATIONS, THEORY OF; POLYNOMIAL SYSTEMS OF EQUATIONS.

Ross A. Beaumont

Disease

A deleterious set of responses which occurs at the subcellular level, stimulated by some injury, and which is often manifested in altered structure or functioning of the affected organism. The concept of disease has been changing as understanding of biological systems has increased. For centuries disease was described in terms of gross changes. Then a major revolution in biology occurred between 1840 and 1860, based largely on the improvement and application of the light microscope. It included the works of Jacob Shleiden and Theodore Schwann, who developed the concept referred to as the cell theory. This theory states that the basic structural and functional unit of all plant and animal life is the cell. R. C. Virchow was the central figure who applied the cell theory to the pathogenesis of human

disease. He recognized that diseases were the product of the alteration of cell structure and function. *See* CELL (BIOLOGY).

A second revolution in biology started in the late 1930s. B. von Borris and E. Ruska in Germany developed the electron microscope, T. Svedberg constructed the ultracentrifuge, and L. Pauling worked on x-ray crystallography. G. W. Beadle and E. Tatum established the gene-enzyme relationship, and J. D. Watson and F. H. Crick described the three-dimensional structure of deoxyribonucleic acid (DNA). With advances in understanding and the development of sensitive probes, it became clear that the fundamental causes of diseases are based on biochemical and biophysical responses within the cell. These responses are still being categorized and, slowly, the mechanisms are being understood. *See* DEOXYRIBONUCLEIC ACID (DNA); ELECTRON MICROSCOPE; ULTRACENTRIFUGE; X-RAY CRYSTALLOGRAPHY.

Homeostasis. Cells are in a constant state of adjustment in response to fluctuating demands and stresses in order to maintain homeostasis. Homeostasis refers to functional equilibrium in an organism and to the processes that maintain it. There is a range of responses that is considered normal. If cells are pushed to respond beyond these limits, there may be an increase, a decrease, or a loss of normal structure or function. These changes may be reversible or irreversible. If irreversible, the cells may die. Thus, subcellular changes may be reflected in altered tissues, organs, and consequently organisms, and results in a condition described as diseased. *See* HOMEOSTASIS.

Manifestations. Lesions are the chemical and structural manifestations of disease. Some may be seen with the naked eye, while others may require microscopic or biochemical examination in order to be recognized. Subjective manifestations of a disease process such as weakness, pain, and fatigue are called symptoms. The objective measurable manifestations such as temperature, blood pressure, and respiratory rate changes are called signs or physical findings. Changes in the chemical or cellular makeup of an organ, tissue, or fluid of the body or its excretory products are called laboratory findings. To make a diagnosis is to determine the nature of the pathologic process by synthesizing information from these sources evaluated in the light of the patient's history and compared with known patterns of signs and symptoms. In common usage, the term disease indicates a constellation of specific signs and symptoms attributable to altered reactions in the individual which are produced by agents that affect the body or its parts. *See* CLINICAL PATHOLOGY.

Etiology. Etiology is the study of the cause or causes of a disease process. Although a disease may have one principal etiologic agent, it is becoming increasingly apparent that there are several factors involved in the initiation of a disease process. Susceptibility of the individual is an ever-present variable.

Endogenous. The etiologic factors can conveniently be divided into two categories. One group consists of endogenous (internal; within the body) factors, and may originate from errors in the genetic material. An example of a disease with an endogenous etiologic factor is sickle-cell anemia. This inherited disorder, occurring almost exclusively in blacks, results in abnormal hemoglobin. There is a mutation, a change in the normal structure of DNA, such that there is a substitution of valine molecule for glutamic acid in one of the polypeptide chains which makes up hemoglobin. It is thought that this results in altered configuration of the hemoglobin molecules, which is reflected in the altered morphology of the red blood cells at low O_2 tension. This change in morphology of red blood cells is referred to as sickling, and when cells assume this shape they tend to have mechanical difficulty in moving through the smaller blood vessels. They also tend to be quite fragile. Sickling may result in thrombosis, anoxia, and anemia. *See* HEMOGLOBIN; HUMAN GENETICS; MUTATION; SICKLE CELL DISEASE.

Exogenous. The other category of etiologic factors is exogenous (environmental). These account for the majority of disease reactions. Exogenous factors include physical, chemical, and biotic agents. The accompanying **table** lists various causative agents and examples of the diseases they produce.

Pathogenesis. Pathogenesis refers to the mechanisms by which the cell, and consequently the body, responds to an etiologic agent. It involves biochemical and physiological responses which are reflected in ultrastructural, microscopic, or gross anatomic lesions. There are a limited number of ways in which cells respond to injury. Thus, lesions produced by several agents may be morphologically similar. The agents may stimulate similar pathogenetic mechanisms. The cellular responses include adaptation, death, disturbances in growth and specialization, mutation, and neoplasia. The nature of the response is modified by the nature of the agent, dose, portal of entry, and duration of exposure, as well as many host factors such as age, sex, nutritional state, and species and individual susceptibility.

Environment. The diseases which are important in causing human death have changed since 1900. At that time, six of the ten leading causes of death in the United States were infectious (biotic) agents. These included tuberculosis, pneumonia, diarrhea, and enteritis and nephritis. During the early phase of the industrial revolution, chemical and physical agents were rarely the cause of disease and were usually limited to a small group of industrial workers. Therefore, control of infectious diseases was the major health care problem. Developments in microbiology fostered better sanitation and public health measures for the control of transmissible organisms. The development of antibiotics and immunization procedures further enabled the control of these agents. At present, only one of the ten leading causes of death in the United States, influenza and pneumonia, is due to biotic agents. *See* ANTIBIOTIC; INDUSTRIAL HEALTH AND SAFETY; INFECTIOUS DISEASE; VACCINATION.

While most of the biotic causes of diseases were being brought under control, continued population

Common exogenous and endogenous causes of disease

Causative agent	Disease
Exogenous factors	
Physical	
Mechanical injury	Abrasion, laceration, fracture
Nonionizing energy	Thermal burns, electric shock, frostbite, sunburn
Ionizing radiation	Radiation syndrome
Chemical	
Metallic poisons	Intoxication from methanol, ethanol, glycol
Nonmetallic inorganic poisons	Intoxication from phosphorus, borate, nitrogen dioxide
Alcohols	Intoxication from methanol, ethanol, glycol
Asphyxiants	Intoxication from carbon monoxide, cyanide
Corrosives	Burns from acids, alkalies, phenols
Pesticides	Poisoning
Medicinals	Barbiturism, salicylism
Warfare agents	Burns from phosgene, mustard gas
Hydrocarbons (some)	Cancer
Nutritional deficiency	
Metals (iron, copper, zinc)	Some anemias
Nonmetals (iodine, fluorine)	Goiter, dental caries
Protein	Kwashiorkor
Vitamins:	
A	Epithelial metaplasia
D	Rickets, osteomalacia
K	Hemorrhage
Thiamine	Beriberi
Niacin	Pellagra
Folic acid	Macrocytic anemia
B_{12}	Pernicious anemia
Ascorbic acid	Scurvy
Biological	
Plants (mushroom, fava beans, marijuana, poison ivy, tobacco opium)	Contact dermatitis, systemic toxins, cancer, hemorrhage
Bacteria	Abscess, scarlet fever, pneumonia, meningitis, typhoid, gonorrhea, food poisoning, cholera, whooping cough, undulant fever, plague, tuberculosis, leprosy, diphtheria, gas gangrene, botulism, anthrax
Spirochetes	Syphilis, yaws, relapsing fever, rat bite fever
Virus	Warts, measles, German measles, smallpox, chickenpox, herpes, roseola, influenza, psittacosis, mumps, viral hepatitis, poliomyelitis, rabies, encephalitis, trachoma
Rickettsia	Spotted fever, typhus
Fungus	Ringworm, thrush, actinomycosis, histoplasmosis, coccidiomycosis
Parasites (animal)	
Protozoa	Amebic dysentery, malaria, toxoplasmosis, trichomonas vaginitis
Helminths (worms)	Hookworm, trichinosis, tapeworm, filariasis, ascariasis
Endogenous factors	
Hereditary	Phenylketonuria, alcaptonuria, glycogen storage disease, Down syndrome (trisomy 21), Turner's syndrome, Klinefelter's syndrome, diabetes, familial polyposis
Hypersensitivity	Asthma, serum sickness, eczema, drug idiosyncrasy

growth (in large part, a consequence of the control of infectious disease) and the remarkable growth of industrialization have been associated with an increased prevalence of diseases caused by physical and chemical agents. Eight of the leading causes of death in the United States have environmental agents as the known or strongly suspected major etiologic factor. These include cancer, cirrhosis, and cardiovascular disease. *See* CANCER (MEDICINE); CIRCULATION DISORDERS; CIRRHOSIS.

General principles of the organism's response to toxic substances, some of which occur naturally in the environment, have evolved from a great number of investigations of agent-host interaction. These principles are: (1) All substances entering the organism are toxic; none is harmless. Dose rate of exposure and route of entry into the body determine whether a toxic response will occur or not. (2) All agents evoke multiple responses. (3) Most of the biological responses are undesirable, leading to the development of pathological changes. (4) A given dose of an agent does not produce the same degree of response in all individuals. Thus, when disease is viewed as interaction between the environment and the individual, the control of disease is largely the management of the environmental causes of disease. *See* CLINICAL MICROBIOLOGY; DEATH; ENVIRONMENTAL TOXICOLOGY; MEDICAL BACTERIOLOGY; PATHOLOGY; TOXICOLOGY; TUMOR.

N. Karle Mottet; Carol Quaife

Geochemical aspects. It is difficult to draw a definite link between geochemical environment and disease except in the few cases where high levels of toxic substances of geochemical origin have contacted organisms. Thus the effects of long-term, low-level exposure are not clearly understood. Only since the 1960s has any systematic and concerted research been carried out to explore the relationship between

geochemical environment and disease. A closely related problem is occupational health as affected by geochemical substances and environmental pollution. *See* INDUSTRIAL HEALTH AND SAFETY.

In 1831 a link was established between the severe disabilities (called miner's disease) shown by miners and the lead metal ores being mined. The Industrial Revolution, and particularly the large demand for metals generated by World Wars I and II, focused great attention on geochemical toxic hazards. The subsequent upsurge in the use of rarer elements in metallurgical and chemical industries has produced a serious health hazard. Very little is known about their environmental and clinical properties.

Of particular interest historically is the effect of geochemical dusts on the lungs of humans. The mining industry was plagued for many years by diseases related to the exposure of miners to high dust levels. Angular dust particles irritate the lining of the lungs and bronchial tubes and, if exposure is high and continuous, destroy the lung tissues, resulting in loss of respiration function. The most common disease of this type is silicosis. The problems caused by siliceous dusts can be largely minimized by using breathing masks, which filter out the offending particles. *See* RESPIRATOR.

Asbestos dusts contain sharp needlelike particles which can easily penetrate lung tissue. Some evidence suggests an abnormally high incidence of lung cancer among workers in the asbestos industry. *See* ASBESTOS.

The toxicity of a metal depends on its physical, chemical, and biochemical properties, the nature of its compounds, its mode of entry into the body, and the individual's susceptibility. The toxic properties of an element are related to its chemical form. For example, swallowing a small drop of metallic mercury could have minimal toxic effect, but a similar quantity of an organomercury compound, such as dimethyl mercury, could be fatal. Also, in the case of chromium, the trivalent form is an essential micronutrient, yet the hexavalent form is carcinogenic. The mode of intake of an element or compound will determine the effectiveness of its assimilation. The factor, together with the rate of excretion, will determine the severity and duration of the harmful effects. Generally, inhalation is a very effective way for ingesting soluble compounds, because these materials transfer directly through the pulmonary tissue into the bloodstream. Insoluble material can also be transported by the cilia of epithelial cells through the pharynx into the intestinal tract. Material injested orally appears in the stomach and is attacked by the digestive acids and enzymes. If it is solubilized in this manner, it can be transported through intestinal cells into body fluids and tissue. Absorption through the skin due to prolonged contact can also result in ingestion into the body. Generally, this is not a very efficient method of assimilation.

The effectiveness of assimilation, and indeed the degree of sensitivity to a toxic material, is dependent on a number of physiological and other factors which may differ from individual to individual. These factors include condition of health, age, body mass, sex, level of stress, and nutritional state. J. C. Van Loon

Disease ecology

The interaction of the behavior and ecology of hosts with the biology of pathogens as relating to the impact of diseases on populations. Although human diseases have been studied from an ecological perspective for centuries, diseases in plant and animal populations have only recently received much attention. The reason that ecologists have generally overlooked diseases is that diseases tend to be "invisible" compared to predators or competitors. The toll that diseases can take on all organisms (even bacteria are attacked by diseases) is enormous, and the ecology of diseases is one of the most vigorous areas of research in modern biology. *See* DISEASE; ECOLOGY.

Threshold theorem. For a disease to spread, on average it must be successfully transmitted to a new host before its current host dies or recovers. This simple observation lies at the core of the most important idea in epidemiology: the threshold theorem. The threshold theorem states that if the density of susceptible hosts is below some critical value, then on average the transmission of a disease will not occur rapidly enough to cause the number of infected individuals to increase. In other words, the reproductive rate of a disease must be greater than 1 for there to be an epidemic, with the reproductive rate being defined as the average number of new infections created per infected individual.

Human immunization programs are based on applying the threshold theorem of epidemiology to public health; specifically, if enough individuals in a population can be vaccinated, then the density of susceptible individuals will be sufficiently lowered that epidemics are prevented. Diseases such as smallpox have been eradicated because vaccination rates were high enough worldwide (70–80% immunization rates) to preclude the possibility of a smallpox outbreak. More generally, because diseases spread rapidly when densities of hosts are high, and less effectively (or not at all) when host densities are low, diseases can profoundly influence the population dynamics of their hosts—either by causing fluctuations in host density or by regulating host densities to low levels in conjunction with other factors. Even human populations seem to have been dramatically influenced by diseases. For example, between the fifth and eleventh centuries, European populations were ravaged by cycles of the plague. The initial plague epidemics killed 25–40% of the urban populations in Europe. Later cycles of plague were less deadly because the populations had evolved resistance and individuals had acquired immunity. *See* EPIDEMIOLOGY; VACCINATION.

Diseases have profoundly altered the course of human history. For example, the conquest of the Aztecs by a few hundred Spanish soldiers was due primarily to the fact that the Spaniards brought to the Western Hemisphere new diseases that decimated

the Aztec populations. In particular, the population of Mexico fell from about 20 million to 3 million between 1518 and 1568. The deaths were from successive epidemics of smallpox, measles, and typhus—diseases that had never before been seen in the New World, but to which the Europeans had built up a resistance (after centuries of exposure). Even today, new diseases can have a major impact on humans. The HIV epidemic in Africa is depleting the population of healthy workers so severely in cities that the economies of many nations are hampered. Almost 20 million people are infected with HIV in sub-Sahara Africa, with many of them being babies.

In general, the rate of reproduction for diseases is proportional to their transmissibility and to the length of time that an individual is infectious. For this reason, extremely deadly diseases that kill their hosts too rapidly may require extremely high densities of hosts before they can spread. All diseases do not behave as simply as hypothesized by the threshold theorem, the most notable exceptions being sexually transmitted diseases. Because organisms actively seek reproduction, the rate at which a sexually transmitted disease is passed among hosts is generally much less dependent on host density. For this reason, sexually transmitted diseases are especially hard to control.

Population effects. Cycles in many animal populations are thought to be driven by diseases. For example, the fluctuations of larch bud moths in Europe are hypothesized to be driven by a virus that infects and kills the caterpillars of this species. Cycles of red grouse in northern England are also thought to be driven by disease, in this case by parasitic nematodes. The interaction between the red grouse and parasitic worms is subtle; the presence of a worm infection does not necessarily mean that a grouse is harmed. Instead, it is only when grouse are laden with heavy worm burdens that effects are seen, and those effects take the form of reduced breeding success or higher mortality during the winter. This example highlights a common feature of diseases: their effects may be obvious only when their hosts are assaulted by other stresses as well (such as harsh winters and starvation). The most severe human plagues in the Middle Ages were associated with times of famine, presumably because malnourished victims were less able to withstand the plague.

Animals. The introduction of novel diseases to wild populations has created massive disruptions of natural ecosystems. For example, the introduction of rinderpest virus into African buffalo and wildebeest populations decimated them in the Serengeti. African wild ungulates have recovered in recent years only because a massive vaccination program eliminated rinderpest from the primary reservoir for the disease, domestic cattle. But the consequences of the rinderpest epidemic among wild ungulates extended well beyond the ungulate populations. For example, human sleeping sickness increased following the rinderpest epidemic because the tsetse flies that transmit sleeping sickness suffered a shortage of game animals (the normal hosts for tsetse flies) and increasingly switched to humans to obtain meals.

Plants. It is widely appreciated that crop plants are attacked by a tremendous diversity of diseases, some of which may ruin an entire year's production. The emigration of Irish people to the United States was prompted by a famine caused by a potato disease. A corn leaf blight destroyed approximately one-third of the United States corn crop in 1970 at a cost of almost $1 billion. Diseases are equally prevalent among wild populations of plants, but their toll seems to be reduced because natural plant populations are so genetically variable that it is unlikely that any given pathogen strain can sweep through and kill all of the plants—there are always some resistant genotypes. But when agronomists have bred plants for uniformity, they have often depleted genetic diversity and created a situation in which a plant pathogen that evolves to attack the crop encounters plants with no resistance (all the plants are the same) and sweeps through the fields with drastic consequences. For example, when leaf blight devastated the United States corn crop, 70% of that crop shared genetically identical cytoplasm, and the genetic uniformity of the host exacerbated the severity of the epidemic. In fact, hundreds of agricultural experiments have been performed in which the severity of plant diseases is contrasted between fields planted with a mixture of varieties and fields planted with only one variety. The results reveal that, on average, genetic variety reduces disease incidence by 50% compared to genetically homogeneous (only one variety) plantings. This is one reason why biodiversity and genetic diversity are viewed as a natural resource; such diversity is likely to harbor resistance traits, which can be bred (or transferred by genetic engineering) into food crops. *See* PLANT PATHOLOGY.

Disease emergence. Humans are dramatically altering habitats and ecosystems. Sometimes these changes can influence disease interactions in surprising ways. Lyme disease in the eastern United States provides a good example of the interplay of human habitat modifications and diseases. Lyme disease involves a spirochete bacterium transmitted to humans by ticks. However, humans are not the normal hosts for this disease; instead, both the ticks and the bacterium are maintained primarily on deer and mouse populations. Human activities influence both deer and mice populations, and in turn tick populations, affecting potential exposure of humans to the disease. The pathways for this influence are numerous and hard to predict. But one well-documented human effect has arisen from the control of natural predators of deer and the clearing of forest habitat; both changes have favored the buildup of large deer populations. As deer populations increase, so do populations of deer ticks and the opportunities for transfer of Lyme disease to humans. Lyme disease could never be maintained in human populations, but excessive deer populations and proximity of deer to human outdoor activities have created a significant health hazard in many northeastern states.

Much less certain are the impacts of anticipated global warming (caused by the burning of fossil

fuels) on diseases. There is clearly some cause for concern about the expansion of tropical diseases into what are now temperate regions in those cases where temperature sets limits to the activity or distribution of major disease vectors (such as the species of mosquito that vector malaria).

The role of natural ecosystems in diseases, and the possibility that human activities can profoundly alter the ecology of diseases, are very significant. *See* LYME DISEASE; POPULATION ECOLOGY. Peter Kareiva

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Dispersion (radiation)

The separation, by refraction, interference, scattering, or diffraction, of acoustic and electromagnetic radiation or energy into its constituent wavelengths or frequencies. For a refracting, transparent substance, such as a prism of glass, the dispersion is characterized by the variation of refractive index with change in wavelength of the radiation. Refractive index (n) is defined as the ratio of the velocity of the radiation in free space (air at standard temperature and pressure for sound, and a vacuum for electromagnetic radiation) to the velocity in the substance in question. I. Newton used a small hole in a window shade and a glass prism (see **illus.**) to disperse sunlight into a visible spectrum, from violet through red. Using a second prism, he showed that no further decomposition of any of the spectral colors could be achieved. *See* OPTICAL PRISM; REFRACTION OF WAVES.

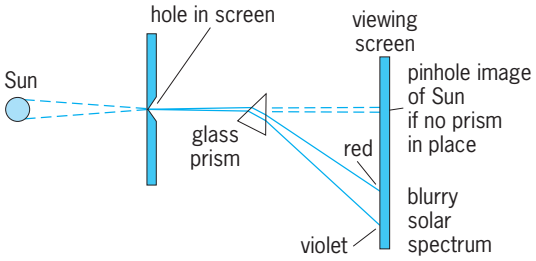
Acoustic radiation can be explained by pure mechanics, but a full understanding of the dispersion of electromagnetic radiation requires atomic theory. *See* ABSORPTION OF ELECTROMAGNETIC RADIATION; SOUND.

Formulas and behavior. The most widely used formula for dispersion is Sellmeier's equation (1), where

$$n^2 - 1 = \sum_{i=0}^{m} \frac{A_i \lambda^2}{\lambda^2 - \lambda_i^2} \tag{1}$$

n is the refractive index at wavelength λ, and A_i, and λ_i, are constants that approximate the asymptotic behavior, bounding but not valid (since n cannot actually go to infinity) in regions of high or total absorption. Sellmeier's equation can be expanded with the binomial theorem, and the dominant terms are those of the older, simpler, but less accurate Cauchy equation (2).

$$n = A + \frac{B}{\lambda^2} + \frac{C}{\lambda^4} \tag{2}$$



Prismatic dispersion of sunlight. (*After J. P. C. Southall, Mirrors, Prisms, and Lenses, Macmillan, 1923*)

Differentiation shows that blue light disperses more rapidly than red light, as the cube of wavelength. Different optical materials have widely varying dispersive power, which is defined by the Abbe number, given by Eq. (3), where d, F, and C refer

$$v_d = \frac{n_d - 1}{n_F - n_C} \tag{3}$$

to Fraunhofer absorption lines in the green, blue, and red parts of the solar spectrum. Different substances disperse wavelengths in slightly different ratios, an optical disadvantage termed irrationality of dispersion, which causes achromatic lenses to have small amounts of residual image color. *See* CHROMATIC ABERRATION; FRAUNHOFER LINES; OPTICAL MATERIALS.

The condition where the refractive index decreases as wavelength increases is termed normal dispersion. The opposite condition is termed anomalous dispersion, and almost always occurs in regions outside the range of visible wavelengths.

Gratings and prisms. Diffraction gratings are flat or curved plates, either reflective or transmissive, with finely ruled lines, that distribute light according to the grating equation (4), where $N =$ the order number, λ is the wavelength, b is the separation between lines ruled onto a plate, and θ_i and θ_d are the angles of incidence and diffraction, relative to the grating normal. Differentiation of θ_d versus λ shows that angular dispersion is almost linear with wavelength. Reflective gratings are usable in virtually any part of the spectrum.

$$N \cdot \lambda = b(\sin\theta_i + \sin\theta_d) \tag{4}$$

Diffraction-based dispersers are normally preferred for measuring the content of the radiation; prism dispersers are generally selected to study material properties of the prism or its contents. Occasionally, gratings and prisms are combined, as in an echelle spectrograph. *See* ASTRONOMICAL SPECTROSCOPY; DIFFRACTION GRATING; SPECTROGRAPH; SPECTROSCOPY. Richard A. Buchroeder

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Dispersion relations

Relations between the real and imaginary parts of a response function. The term dispersion refers to the fact that the index of refraction of a medium is a function of frequency. In 1926 H. A. Kramers and R. Kronig showed that the imaginary part of an index of refraction (that is, the absorptivity) determines the real part (that is, the refractivity); this is called the Kramers-Kronig relation. The term dispersion relation is now used for the analogous relation between the real and imaginary parts of any response function, such as Eq. (14) below.

Response function. Consider a system in which a cause $C(t)$ [for example, a force] and its effect $E(t)$ [for example, a displacement] are related by Eq. (1),

$$E(t) = \int_{-\infty}^{t} dt'\, G(t-t')C(t') \tag{1}$$

where $G(t)$ is called the response function. Because this relation is linear in both $C(t)$ and $E(t)$, the response is said to be linear: The superposition of two causes results in the sum of their effects. Because the response function G is a function only of $t - t'$, the response is said to be time-independent: The effect at time t of a cause at time t' depends only on the time difference $t - t'$. Because the upper limit of integration is t [or equivalently, $G(t)$ can be said to vanish for negative argument, Eq. (2)], the response

$$G(t) = 0 \qquad \text{if } t < 0 \tag{2}$$

is said to be causal: the cause has no effect at earlier times. *See* CAUSALITY; LINEARITY.

Many examples can be given. A few pairs of cause and effect are: force and spatial displacement, electric field and polarization (G is electrical susceptibility), and incident wave and scattered wave (G is scattering amplitude).

Pure-tone response. The Fourier transform $g(\omega)$ of the response function $G(t)$ is defined in Eq. (3) and the inverse relation in Eq. (4). The function $g(\omega)$

$$g(\omega) = \int_{0}^{\infty} dt\, e^{i\omega t} G(t) \tag{3}$$

$$G(t) = \frac{1}{2\pi} \int_{-\infty}^{\infty} d\omega\, e^{-i\omega t} g(\omega) \tag{4}$$

is complex-valued; however, if $G(t)$ is real [which it must be if $C(t)$ and $E(t)$ of Eq. (1) are real], the crossing relation, Eq. (5), is valid, where an asterisk

$$g^*(\omega) = \int_{0}^{\infty} dt\, e^{-i\omega^* t} G(t) = g(-\omega^*) \tag{5}$$

indicates complex conjugation. The function $g(\omega)$ is called the frequency-dependent response; it describes the response to a simple harmonic (pure tone) cause: if the cause is given by Eq. (6), then

$$C(t) = \cos(\omega t - \eta),\ \omega \text{ real} \tag{6}$$

Eq. (1) implies that its effect is given by Eq. (7),

$$\begin{aligned} E(t) &= \int_{-\infty}^{\infty} dt'\, G(t-t')\cos(\omega t' - \eta) \\ &= {}^1\!/_2 \int_{-\infty}^{\infty} dt'\, G(t-t') e^{-i(\omega t' - \eta)} + \text{c.c.} \\ &= {}^1\!/_2 e^{-i(\omega t - \eta)} \int_{-\infty}^{\infty} dt'\, G(t-t') e^{i\omega(t-t')} + \text{c.c.} \\ &= {}^1\!/_2 e^{-i(\omega t - \eta)} g(\omega) + \text{c.c.} \\ &= {}^1\!/_2 \left|g(\omega)\right| e^{-i(\omega t - \eta - \delta(\omega))} + \text{c.c.} \\ &= \left|g(\omega)\right| \cos\left[\omega t - \eta - \delta(\omega)\right] \end{aligned} \tag{7}$$

where c.c. means complex conjugate and $|g|$ and δ are given by Eq. (8). One sees that the magnitude $|g|$

$$g(\omega) = \left|g(\omega)\right| e^{i\delta(\omega)} \tag{8}$$

and phase δ of $g(\omega)$ are the amplitude and phase shift, respectively, of the simple harmonic effect relative to the cause. *See* FOURIER SERIES AND TRANSFORMS; INTEGRAL TRANSFORM.

Causality. Causality, Eq. (2), has an important consequence for $g(\omega)$, namely the dispersion relations given in Eq. (14) below. Their derivation will be outlined, using Cauchy's formula and other properties of analytic functions of a complex variable. Let ω take on complex values, $\omega = \text{Re}\ \omega + i\ \text{Im}\ \omega$. Then $g(\omega)$ as given by Eq. (3) is an analytic function of ω as long as the integral, and its derivative with respect to ω, exist (that is, are finite). For any stable system, $G(t)$ is bounded in magnitude for all values of t. It follows that the integral in Eq. (3) exists for all ω such that $\text{Im}\ \omega > 0$ ("in the upper half ω-plane") because then the magnitude of the factor $e^{i\omega t}$ of the integrand falls exponentially with increasing t according to Eq. (9).

$$\left|e^{i\omega t}\right| = e^{-(\text{Im}\ \omega)t} \tag{9}$$

Since, on the other hand, for decreasing t this factor rises exponentially, the existence of the integral depends on causality, Eq. (2). So, causality implies that $g(\omega)$ is an analytic function in the upper half ω-plane. *See* COMPLEX NUMBERS AND COMPLEX VARIABLES.

Consequently one may write Cauchy's formula for $g(\omega)$, Eq. (10), where the integration contour C is

$$g(\omega) = \frac{1}{2\pi i} \int_{C} d\omega' \frac{g(\omega')}{\omega' - \omega} \tag{10}$$

any simple loop, traversed counterclockwise, which surrounds the point ω and which lies in the upper half ω'-plane [the region where $g(\omega)$ is an analytic function]. Letting the point ω approach the contour C, the Cauchy formula comes to the form Eq. (11),

$$g(\omega) = \frac{1}{2\pi i} \left[\text{P} \int_{C} \frac{d\omega'}{\omega' - \omega} g(\omega') + i\pi g(\omega) \right] \tag{11}$$

where P means principal value. [Roughly speaking, it means: omit from the range of integration the interval of C which lies within a distance ϵ on either side of the singular point $\omega' = \omega$, and then take the limit

$\epsilon \to 0$.] Solving for $g(\omega)$ gives Eq. (12), where HT

$$g(\omega) = \frac{-i}{\pi} \mathrm{P} \int_C \frac{d\omega'}{\omega' - \omega} g(\omega) \equiv -i\mathrm{HT}[g] \qquad (12)$$

means Hilbert transform (using the contour C). Now let the contour C be expanded until it encloses the entire upper half ω' plane. If $g(\omega)$ vanishes at large ω, the only part of the contour which contributes to the Hilbert transform integral runs along the real ω' axis. Taking also ω to be real, the real and imaginary parts of Eq. (12) give Eqs. (13), where Re g and

$$\mathrm{Re}\, g = \mathrm{HT}[\mathrm{Im}\, g] \qquad \mathrm{Im}\, g = -\mathrm{HT}[\mathrm{Re}\, g] \qquad (13)$$

Im g are the real and imaginary parts of g. Finally, using crossing, Eq. (5), the integral can be expressed as an integral over just positive ω', Eqs. (14).

$$\begin{aligned} \mathrm{Re}\, g(\omega) &= \frac{2}{\pi} \mathrm{P} \int_0^\infty d\omega' \, \frac{\omega' \, \mathrm{Im}\, g(\omega')}{\omega'^2 - \omega^2} \\ \mathrm{Im}\, g(\omega) &= \frac{-2\omega}{\pi} \mathrm{P} \int_0^\infty d\omega' \, \frac{\mathrm{Re}\, g(\omega')}{\omega'^2 - \omega^2} \end{aligned} \qquad (14)$$

These are the dispersion relations for the frequency-dependent response function $g(\omega)$. They say that Re g and Im g are not independent functions; given one of them, the other is determined.

Resonance. Although a causal $g(\omega)$ is analytic in the upper ω plane, it can, and usually does, have singularities in the lower ω-plane. For instance, $g(\omega)$ may have a pole at $\omega = \omega_0$ [and to satisfy Eq. (5) likewise at $\omega = -\omega_0{}^*$]; causality requires that Im $\omega_0 < 0$. In Eq. (15) such a pole term of $g(\omega)$ is

$$\begin{aligned} g(w) &= \frac{A}{\omega - \omega_0} + \frac{A}{-\omega - \omega_0^*} + \cdots \\ &= \frac{\mathrm{Re}\, A + i\, \mathrm{Im}\, A}{\omega - \omega_R + i\, {}^1\!/_2\gamma} + \frac{\mathrm{Re}\, A - i\, \mathrm{Im}\, A}{-\omega - \omega_R - i\, {}^1\!/_2\gamma} + \cdots \end{aligned} \qquad (15)$$

written out, with the notation Re $\omega_0 = \omega_R$, Im $\omega_0 = -{}^1\!/_2\gamma$. If γ is small, then $|g(\omega)|^2$ as a function of real ω peaks sharply at $\omega = \omega_R$ with a width (defined as the interval of ω between the half height points, where $|g(\omega)|^2$ is half as big as the peak) equal to γ. This is called a resonance of the response function $g(\omega)$; ω_R is its position and γ is its width. *See* RESONANCE (ACOUSTICS AND MECHANICS); RESONANCE (ALTERNATING-CURRENT CIRCUITS).

The consequence of a resonance for the effect $E(t)$ is found by using Eq. (15) in Eqs. (1) and (4). If the cause $C(t)$ turns off rapidly enough, then at large t Eq. (16) is obtained, where the ω integration was done by residues. The result is that the effect $E(t)$ has an exponentially decaying term whose time constant (for E^2) is γ^{-1}; that is, the width of the resonance peak of $|g(\omega)|^2$ and the decay constant of $E(t)^2$ are reciprocals of one another. In more physical terms, the decaying oscillation of $E(t)$ reflects a (decaying) normal mode of oscillation, a "ringing" of the system.

$$\begin{aligned} E(t) &= \int dt' \, C(t') \frac{1}{2\pi} \\ &\quad \cdot \int d\omega \, e^{-i\omega(t-t')} \frac{A}{\omega - \omega_0} + \text{c.c.} + \cdots \xrightarrow{t\to\infty} \\ &\quad -iA \int dt' \, C(t') e^{-i\omega_0(t-t')} + \text{c.c.} + \cdots \\ &= \left[-iAe^{-i\omega_R t} \int dt' \, C(t') e^{i\omega_0 t'} + \text{c.c.} \right] e^{\gamma t/2} + \cdots \\ &= C_1 \cos(\omega_R t + C_2) \, e^{-\gamma t/2} + \cdots \end{aligned} \qquad (16)$$

(where C_1, C_2 are constants)

If Im ω_0 were positive, contrary to causality, the calculation of Eqs. (16) would show the effect $E(t)$ to have an exponentially growing term at an early time, preceding the cause. As for $g(\omega)$, the qualitative effect of causality is that at the resonance, where the magnitude $|g(\omega)|$ has a narrow peak, the phase of $g(\omega)$ [$\delta(\omega)$ in Eq. (7)] rises rapidly; if Im ω_0 were positive, contrary to causality, the phase would instead fall rapidly.

Dielectric constant and refractive index. The response of a macroscopically homogeneous dielectric (polarizable) medium to an electromagnetic plane wave of frequency ω is described by the dielectric constants $\epsilon_\parallel(\omega)$ and $\epsilon_\perp(\omega)$. The subscripts distinguish whether the electric field is parallel or perpendicular to the direction of propagation of the plane wave. The discussion will be limited to $\epsilon_\perp(\omega)$, which will be referred to as simply $\epsilon(\omega)$. The refractive index $n(\omega)$ is simply related: $n^2 = \epsilon$. But $n(\omega)$ itself will not in general satisfy dispersion relations because if $\epsilon(\omega)$ has a zero in the upper ω plane, $\sqrt{\epsilon} = n$ will have a square root singularity. *See* PERMITTIVITY.

At $\omega = \infty$, $\epsilon(\omega)$ has the value 1 (the vacuum value). At $\omega = 0$, $\epsilon(\omega)$ is singular if the medium is conductive (that is, contains charges which are free to move), according to Eq. (17), where σ_0 is the static conductivity and ω_0 is the plasma frequency; n_v is the number density of free charges of charge e and mass m. Thus the function given in Eq. (18) has the properties assumed for $g(\omega)$ above and can be substituted for it in Eq. (14), resulting in the dispersion relations, Eqs. (19). In Eqs. (19), terms have been added to the integrands in order to make them nonsingular; these terms do not alter the integrals because $\mathrm{P} \int d\omega'/(\omega'^2 - \omega^2) = 0$.

$$\epsilon(\omega) \xrightarrow{\omega\to 0} \frac{4\pi i\sigma_0}{\omega} - \frac{\omega_0^2}{\omega^2}, \quad \omega_0^2 = \frac{n_v 4\pi e^2}{m} \qquad (17)$$

$$g(\omega) = \epsilon(\omega) - 1 - \frac{4\pi i\sigma_0}{\omega} + \frac{\omega_0}{\omega^2} \qquad (18)$$

$$\begin{aligned} \mathrm{Re}\, \epsilon(\omega) &= 1 - \frac{\omega_0^2}{\omega^2} + \frac{2}{\pi} \int_0^\infty d\omega' \\ &\quad \cdot \frac{\omega' \, \mathrm{Im}\, \epsilon(\omega') - \omega \, \mathrm{Im}\, \epsilon(\omega)}{\omega'^2 - \omega^2} \end{aligned} \qquad (19a)$$

$$\begin{aligned} \mathrm{Im}\, \epsilon(\omega) &= \frac{4\pi\sigma_0}{\omega} - \frac{2}{\pi} \omega \int_0^\infty d\omega' \\ &\quad \cdot \frac{\mathrm{Re}\, \epsilon(\omega') + \omega_0^2/\omega'^2 - \mathrm{Re}\, \epsilon(\omega) - \omega_0^2/\omega^2}{\omega'^2 - \omega^2} \end{aligned} \qquad (19b)$$

By setting ω equal to special values, interesting results are obtained; for simplicity these will be given for the case of a nonconductor, where $\sigma_0 = \omega_0 = 0$. Setting $\omega = 0$ in Eq. (19a) results in an expression

for the static dielectric constant, Eq. (20). which are

$$\epsilon(0) = 1 + \frac{2}{\pi}\int_0^\infty d\omega\ \frac{\operatorname{Im}\epsilon(\omega)}{\omega} \tag{20}$$

$$\operatorname{Re}\epsilon(\omega) \to 1 - \frac{\omega_\infty^2}{\omega^2} \tag{21a}$$

where $$\omega_\infty^2 = \frac{2}{\pi}\int_0^\infty d\omega\ \omega\ \operatorname{Im}\epsilon(\omega)$$

$$\operatorname{Im}\epsilon(\omega) \to \frac{4\pi\sigma_\infty}{\omega} \tag{21b}$$

where $$4\pi\sigma_\infty = \frac{2}{\pi}\int_0^\infty d\omega\,[\operatorname{Re}\epsilon(\omega) - 1]$$

sum rules. The first of these relates an integral over all frequencies of the absorptivity, Im ϵ, to the quantity ω_∞^2, for which a direct calculation gives the value $\omega_\infty^2 = \Sigma\ n_v 4\pi e^2/m$, where the sum is over all kinds of charges in the medium, in contrast to the formula for the plasma frequency ω_0, Eq. (17); at infinite frequency all particles of the medium behave as though free. *See* ABSORPTION OF ELECTROMAGNETIC RADIATION.

Forward-scattering amplitude. It can be shown by arguments like those used above that the forward-scattering amplitude $f(\omega)$ of light incident on an arbitrary body satisfies the dispersion relation given in Eq. (22) [the less interesting relation giving Im f

$$\operatorname{Re} f(\omega) = f(0) + \frac{\omega^2}{2\pi^2 c}\int_0^\infty d\omega'\ \frac{\sigma_T(\omega') - \sigma_T(\omega)}{\omega'^2 - \omega^2} \tag{22}$$

where $$f(0) = \frac{-Q^2}{Mc^2}$$

in terms of Re f has been omitted], where Q is the charge, M is the mass of the body, and c is the speed of light. In Eq. (22) Im $f(\omega)$ has been expressed in terms of the total cross section σ_T by the optical theorem, Eq. (23); this formula is true for the scattering

$$\operatorname{Im} f(\omega) = \frac{k}{4\pi}\sigma_T(\omega) \tag{23}$$

where k = wave number

of waves of any sort.

Equations (22) and (14) can be related by using the "optical potential" formula, Eq. (24), which relates

$$\frac{k^2(\text{in medium})}{k^2(\text{in vacuum})} = n^2 - 1 + \frac{4\pi n_v f(\omega)}{k^2} \tag{24}$$

the forward scattering amplitude $f(\omega)$, of a wave (of any sort) of frequency ω and wave number k incident on a body, to the index of refraction n of that wave in a medium consisting of a random arrangement of such bodies with mean number density n_v. However, Eq. (24) is valid only for a dilute medium, whereas Eqs. (14) and (22) are both exact if causality is true.

An interesting consequence of Eq. (22) is that a system consisting of the electromagnetic field and point charges, obeying classical mechanics, cannot be causal. For such a system it can be shown that $f(\omega)$ vanishes for $\omega = \infty$; using this in Eq. (22) gives Eq. (25), which is untrue since both terms on the

$$0 = f(0) - (2\pi^2 c)^{-1}\int_0^\infty d\omega\sigma_T \tag{25}$$

right-hand side are negative. In the quantum theory, by contrast, $\sigma_T > 0(1/\omega)$ and so no formula like Eq. (25) follows from causality. *See* NONRELATIVISTIC QUANTUM THEORY.

Elementary particle theory. In the quantum theory the scattering amplitudes of matter waves satisfy dispersion relations similar to Eq. (22) if causality holds. This gives the possibility of experimentally testing causality up to very high frequencies (corresponding to very short time intervals) by using, for instance, protons of energy up to 1000 GeV, incident on protons in a target, at Fermilab. Since the right-hand side of Eq. (22) requires knowledge of σ_T at all ω, one can strictly say only that the observed $f(\omega)$ is consistent with causality and a reasonable extrapolation of σ_T beyond observed energies.

Not only is there no experimental evidence against causality, but causality has always been a property of relativistic quantum field theory (elementary particle theory). The following is essentially equivalent to causality: Considered as functions of the Lorentz invariant scalars formed from the 4-momenta of the interacting particles (for instance, the total energy and the momentum transfer in the center-of-mass frame, in the case of a two-body scattering amplitude), the S-matrix elements (scattering amplitudes) of quantum field theory are analytic functions. More precisely, they are analytic except for singularities on Landau surfaces in the space of the (complex) arguments, the Lorentz scalars. These singularities are branch points, and if the amplitude is analytically continued around one, the amplitude becomes changed by a certain amount.

On the one hand, knowing this change allows the amplitude itself to be determined, by use of Cauchy's formula [a dispersion relation like Eq. (22) is a simple example of this]. On the other hand, the change is given in terms of other S-matrix elements. [The optical theorem, Eq. (23), is a simple example of this; the point is that σ_T is a sum (and integral) over products of two amplitudes.] These S-matrix equations, resulting from the combination of the analyticity and the unitarity of the S-matrix, are the subject of S-matrix theory. *See* RELATIVITY.

These equations can to some extent replace the equations of motion of quantum field theory. They have the advantage over the only known systematic solution of the equations of motion, namely perturbation theory (Feynman diagrams), in that they involve only S-matrix elements, that is, the amplitudes of real, observable processes. However, there is no known way to find the solution of the S-matrix equations which corresponds to a given set of field theory equations of motion.

It was the original hope of many workers in S-matrix theory that the equations of motion were in fact irrelevant and that the S-matrix was determined

by a few properties of low-mass particles (in fact perhaps by none, in the extreme view termed bootstrapping). However, the successes of quantum chromodynamics (QCD) in describing and predicting many phenomena in high-energy physics seems to show that a specific field theory, whose elementary quanta (quarks and gluons) are not observable as particles, underlies the S-matrix of "elementary" particle interactions. The S-matrix equations are true in any case, and will remain an important tool in the description and correlation of experimental data. *See* ELEMENTARY PARTICLE; QUANTUM CHROMODYNAMICS; QUANTUM FIELD THEORY; SCATTERING EXPERIMENTS (NUCLEI); SCATTERING MATRIX. Charles J. Goebel

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Displacement current

The name given by J. C. Maxwell to the term $\partial \mathbf{D}/\partial t$ which must be added to the current density $\mathbf{i}$ to extend to time-varying fields A. M. Ampère's magnetostatic result that $\mathbf{i}$ equals the curl of the magnetic intensity $\mathbf{H}$. In integral form this result is given by Eq. (1), where the unit vector $\mathbf{n}$ is perpendicular to

$$\oint \mathbf{H} \cdot d\mathbf{s} = \int_S \left(\mathbf{i} + \frac{\partial \mathbf{D}}{\partial t} \right) \cdot \mathbf{n}\, dS \qquad (1)$$

the surface dS. The concept of displacement current has important consequences for insulators and for free space where $\mathbf{i}$ vanishes. For conductors, however, the difference between Eq. (1) and Ampère's result is negligible. *See* AMPÈRE'S LAW; MAXWELL'S EQUATIONS.

In order to show that the displacement term is essential, consider a parallel-plate capacitor charged by a circuit carrying an alternating current. Let a closed curve s encircle one of the charging wires and be the boundary of two surfaces, S_1 which passes through the capacitor gap and S_2 which cuts the charging wire. By Gauss' electric flux theorem, the charge Q on the plate and wire between S_1 and S_2 is given by Eq. (2), where the normal is taken in the direction

$$Q = \int_{S_1} \mathbf{D} \cdot \mathbf{n}\, dS - \int_{S_2} \mathbf{D} \cdot \mathbf{n}\, dS \qquad (2)$$

of current flow in both S_1 and S_2. The current $\mathbf{I}$ or $\int \mathbf{i} \cdot \mathbf{n} dS$ equals $\partial Q/\partial t$ so that Eq. (3) holds. Thus, the

$$\int_{S_1} \frac{\partial \mathbf{D}}{\partial t} \cdot \mathbf{n}\, dS = \int_{S_2} \left(\mathbf{i} + \frac{\partial \mathbf{D}}{\partial t} \right) \cdot \mathbf{n}\, dS$$

$$= \oint \mathbf{H} \cdot d\mathbf{s} \qquad (3)$$

inclusion of the displacement current is needed to make Eq. (1) valid for any surface S bounded by s.

If one defines current as a transport of charge, the term displacement current is a misnomer when applied to a vacuum where no charges exist. If, however, current is defined in terms of the magnetic fields it produces, the expression is legitimate. In a dielectric, where an electric field produces a displacement of the negative charges with respect to the positive ones, the name has meaning. Maxwell had this sort of picture, even for a vacuum, where he postulated a polarizable ether. *See* ETHER HYPOTHESIS. William R. Smythe

Displacement pump

A pump that develops its action through the alternate filling and emptying of an enclosed volume.

Reciprocating types. Positive-displacement reciprocating pumps have cylinders and plungers or pistons with an inlet valve, which opens the cylinder to the inlet pipe during the suction stroke, and an outlet valve, which opens to the discharge pipe during the discharge stroke. Reciprocating pumps may be power-driven through a crank and connecting rod or equivalent mechanism, or direct-acting, driven by steam or compressed air or gas.

Figure 1 shows a small high-speed plunger-type power pump for high-pressure service. The three-throw crankshaft is carried in roller bearings at each end. The manifolds below the suction valves and above the discharge valves connect the three pumping cylinders to the suction and discharge piping.

Power pumps are frequently built with one or two throw cranks and double-acting liquid ends, or with five, seven, or even nine cranks where smoother flow

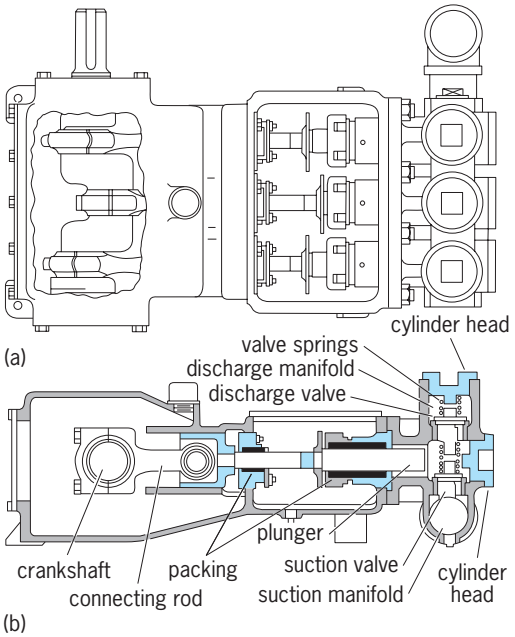

Fig. 1. Horizontal triplex power pump of reciprocating type. (*a*) Plan. (*b*) Elevation.

is desirable. Power-driven reciprocating pumps are highly efficient over a wide range of discharge pressures. Except for some special designs with continuously variable stroke, reciprocating power pumps deliver essentially constant capacity over their entire pressure range when driven at constant speed. In some applications this is an advantage, but in others it complicates the controls required.

Direct-acting steam types. A reciprocating pump is readily driven by a reciprocating engine; a steam or power piston at one end connects directly to a fluid piston or plunger at the other end. Direct-acting reciprocating pumps are simple, flexible, low-speed machines which are low in efficiency unless the heat in the exhaust steam can be used for heating. Steam pumps can be built for a wide range of pressure and capacity by varying the relative size of the steam piston and the liquid piston or plunger. The delivery of a steam pump may be varied at will from zero to maximum simply by throttling the motive steam, either manually or by automatic control. Direct-acting pumps are built as simplex, having one steam and one liquid cylinder; and duplex, having two steam and two liquid cylinders side by side. As indicated by **Fig. 2**, each steam valve of a duplex pump is positively driven by the motion of the opposite piston rod by means of cranks and links. In the case of a simplex pump, to avoid stalling at low speed, the valve linkage operates a small pilot valve which in turn controls a piston-thrown main steam valve.

Reciprocating pumps are used for low to medium capacities and medium to highest pressures. They are useful for low- to medium-viscosity fluids, or high-viscosity fluids at materially reduced speeds. Specially fitted reciprocating pumps are used to pump fluids containing the more abrasive solids.

Rotary types. Another form of displacement pump consists of a fixed casing containing gears, cams, screws, vanes, plungers, or similar elements actuated by rotation of the drive shaft. Most forms of rotary pumps are valveless and develop an almost steady flow rather than the pulsating flow of a reciprocating pump. Three of the many types of rotary pumps are shown in **Fig. 3**.

Rotary pumps require very close clearances between rubbing surfaces for their continued volumetric efficiency. Consequently they are most useful for pumping clean oils or other fluids having lubricating value and sufficient viscosity to prevent excessive leakage. On petroleum oils of suitable viscosity, rotary pumps are highly efficient at moderate pressure and speed, while at reduced speed they can pump, with lower efficiency, the most viscous materials. The increasing use of hydraulic actuation of machine tools and mechanisms, such as power steering of automobiles, has extended the use of rotary pumps and similar hydraulic motors.

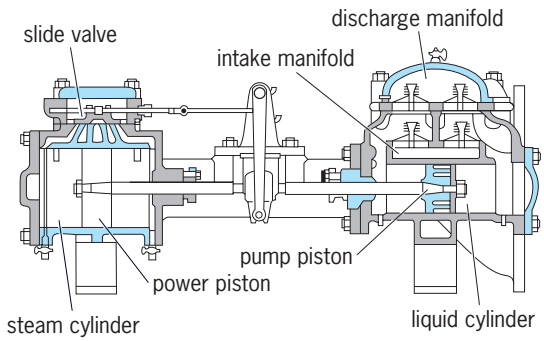

Fig. 2. Duplex type of direct-acting steam-driven feedwater pump.

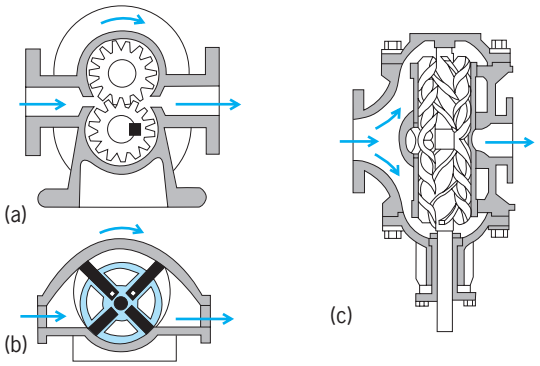

Fig. 3. Rotary pumps. (*a*) Gear. (*b*) Guided vane. (*c*) Screw. (*After E. A. Avallone and T. Baumeister III, eds., Marks' Standard Handbook for Mechanical Engineers, 9th ed., McGraw-Hill, 1987*)

Vacuum types. Although vacuum pumps actually function as compressors, displacement pumps are used for certain vacuum pump applications. Simplex steam pumps with submerged piston pattern fluid ends similar to the fluid end in Fig. 2 are used as wet vacuum pumps in steam heating and condensing systems. Sufficient liquid remains in the cylinder to fill the clearance volume and drive the air or gas out ahead of the liquid. Certain types of rotary pumps are arranged to retain a quantity of sealing oil when operating as vacuum pumps. *See* VACUUM PUMP.

Air-lift types. In handling abrasive or corrosive waters or sludges, where low efficiency is of secondary importance, air-lift pumps are used. The pump consists of a drop pipe in a well with its lower end submerged and a second pipe which introduces compressed air near the bottom of the drop pipe. The required submergence varies from about four times the distance from the water level to the surface for a low lift to an equal distance for a relatively high lift. The mixture of air and water in the drop pipe is lighter than the water surrounding the pipe. As a result, the mixture of air and water is forced to the surface by the pressure of submergence. *See* COMPRESSOR; PUMP; PUMPING MACHINERY. Elliott E. Wright

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Distance-measuring equipment

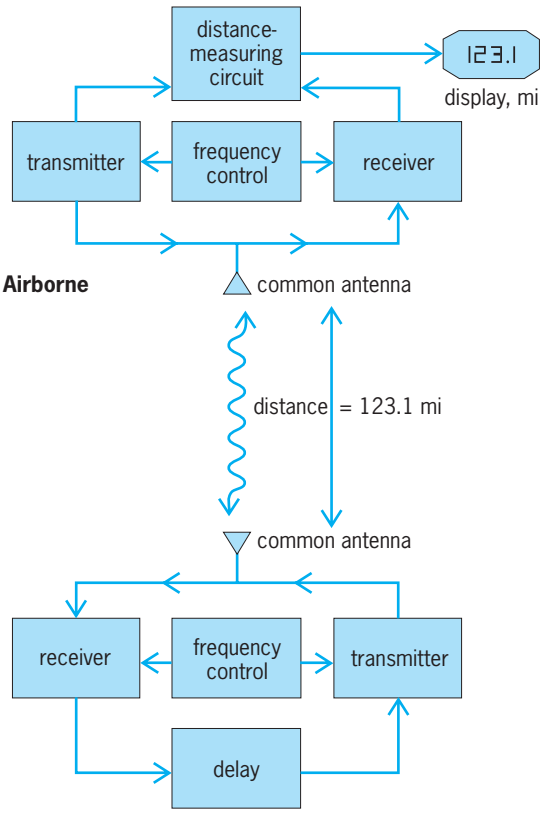
An internationally standardized navigation system which allows an aircraft to measure its distance from a selected ground-based beacon. Such beacons are used throughout the world by all airliners, most of

the military aircraft of the West, and a large number of general-aviation aircraft. The range of service is line of sight up to 300 mi (480 km), and system accuracy is usually 0.1 mi (0.16 km), but precision equipment, intended for use during landing, has accuracy of 20 ft (7 m).

The airborne equipment, called an interrogator, transmits pulses of 1 kW peak power on 1 of 126 frequencies. These are in the 1025–1150-MHz band and are spaced 1 MHz apart. Each pulse has a duration of 3.5 microseconds and is paired with another, spaced 12 or 36 μs later. The combination of frequencies and pulse spacings therefore provides 252 operating channels.

The beacon on the ground, called a transponder, receives these pulses, delays them by 50 μs, and then retransmits them, usually with a power of 1 kW, on 252 frequencies lying between 962 and 1213 MHz. The pulse-pair spacing is 12 μs on those frequencies not used by the interrogator, and 30 μs on those frequencies shared with the interrogator. The transponder transmission is called the reply. The frequency difference between interrogation and reply is always 63 MHz. This arrangement allows each transmitter frequency to act as the local oscillator for its associated superheterodyne receiver, the intermediate frequency of which is 63 MHz. For landing purposes, some transponders have powers as low as 100 W (see **illus.**).

The airborne transmission has a pseudorandom pulse repetition rate ranging from about 15 pulse pairs per second in the newest equipment to as high as 150 pulse pairs per second in older equipment. This reduction is effected to accommodate a larger number of aircraft per beacon. With an average rate of 30 interrogations per second, a typical beacon can reply to 100 aircraft.



Block diagram of distance-measuring equipment system.

In the aircraft the replies to its own interrogations are recognized by their phase coherence with their own transmissions, and by the elapsed time measured between transmissions and reception (minus the 50-μs transponder delay), usually by means of a crystal clock. This elapsed time is about 12 μs for each nautical mile (7 μs for each kilometer), and is displayed in the cockpit on a digital meter, which is usually calibrated in miles and tenths of miles.

Transmission, both airborne and ground, is vertically polarized and employs the same antenna as for reception. The airborne antenna is typically a 5-in. (13-cm) blade (half wavelength) protruding from the belly of the aircraft. The ground antenna usually employs several half-wave sections, totaling about 5 ft (1.5 m) in height.

Much airborne equipment, particularly that used by general aviation, has small adapter units which compute radial speed with respect to the beacon or time-to-go with respect to the beacon, or both. By comparing these numbers with the indicated airspeed, the magnitude of headwind may be gaged. This factor is of greater importance to small aircraft than to airliners.

Typical airliner equipment weighs 12 lb (5.4 kg), while general-aviation types are about 5 lb (2.3 kg). All are entirely solid-state.

Of the 252 two-way channels, 200 are permanently paired on a one-for-one basis with the 200 channels used by the very high-frequency omnidirectional range (VOR) and its associated instrument landing system (ILS) localizer. Thus, in all civil aircraft, a common single channel selector is used for distance-measuring equipment (DME), VOR, and ILS. This rho-theta navigation system is the backbone of the airways throughout the world. It is fully implemented in the United States, Europe, and Japan, with somewhat less implementation elsewhere. A notable latter exception is Australia, where the DME, while also used with VOR, operates in the 200-MHz band. Tacan is a military alternative rho-theta system based on DME.

Jet-powered airliners, typically flying above 15,000 ft (4500 m), are frequently within range, simultaneously, of many DME beacons, whose geographical spacings have been based on service to lower-flying aircraft. This has led to the airborne scanning-DME which automatically scans all beacons within range and then selects two or more to provide a rho-rho (DME/DME) fix of somewhat greater accuracy than that obtained from rho-theta (DME/VOR).

A modification called DME-P, for precision, has been developed for use with the microwave landing system (MLS). This system can provide an accuracy of 20 ft (7 m) within 7 mi (11 km) and 100 ft (30 m) within 20 mi (12 km), using the same frequencies as the standard system (now called

DME-N), but with steeper pulse rise times, lower power, and new pulse-pair codes. *See* AIR NAVIGATION; ELECTRONIC NAVIGATION SYSTEMS; INSTRUMENT LANDING SYSTEM (ILS); MICROWAVE LANDING SYSTEM (MLS); RHO-THETA SYSTEM; TACAN; VOR (VHF OMNIDIRECTIONAL RANGE). Sven H. Dodington

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Distillation

A method for separating homogeneous mixtures based upon equilibration of liquid and vapor phases. Substances that differ in volatility appear in different proportions in vapor and liquid phases at equilibrium with one another. Thus, vaporizing part of a volatile liquid produces vapor and liquid products that differ in composition. This outcome constitutes a separation among the components in the original liquid. Through appropriate configurations of repeated vapor-liquid contactings, the degree of separation among components differing in volatility can be increased manyfold. *See* PHASE EQUILIBRIUM.

Distillation is by far the most common method of separation in the petroleum, natural gas, and petrochemical industries. Its many applications in other industries include air fractionation, solvent recovery and recycling, separation of light isotopes such as hydrogen and deuterium, and production of alcoholic beverages, flavors, fatty acids, and food oils.

Simple distillations. The two most elementary forms of distillation are a continuous equilibrium distillation and a simple batch distillation (**Fig. 1**).

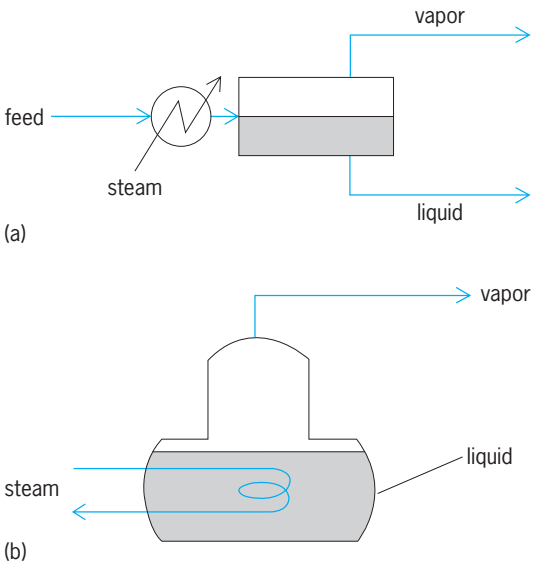

Fig. 1. Simple distillations. (*a*) Continuous equilibrium distillation. (*b*) Simple batch distillation.

In a continuous equilibrium distillation, a continuously flowing liquid feed is heated or reduced in pressure (flashed) so as to cause partial vaporization. The vapor and liquid disengage while flowing through an open drum, and the products emerge as vapor and liquid streams. The vapor product can be condensed to form a liquid distillate. It is also possible to use a vapor feed, subjected to cooling and thereby partial condensation, again followed by disengagement of the resultant vapor and liquid in an open drum. *See* VAPOR CONDENSER.

In a simple batch distillation, an entire batch of liquid is initially charged to a vessel and is then heated, typically by condensation of steam inside a metal coil within the vessel. Vapor is thereby continuously generated, and may be condensed to form a liquid distillate, which is collected. In the batch distillation, increments of vapor are formed in equilibrium with all liquid compositions ranging from the original to the final, whereas the continuous equilibrium distillation gives vapor in equilibrium with only the final liquid composition. Since the distillate consists primarily of the more volatile components and the feed liquid contains more of these substances than does the final liquid, the simple batch distillation gives a more enriched distillate than does the continuous equilibrium distillation.

Fractional distillation. Unless the vapor pressures of the species being separated are very dissimilar, a simple distillation does not produce highly purified products. Product purities can be increased by repeated partial vaporizations and condensations (**Fig. 2***a*). The liquid from an initial continuous equilibrium distillation (L_0') can be partially vaporized by additional heating. The remaining liquid (L_1') can again be heated and partially vaporized, forming another liquid (L_2'); and so forth. Each liquid is progressively enriched in the less volatile substances. Similarly, successive partial condensations of the vapor fraction (V) from the initial continuous equilibrium distillation produce vapor products successively enriched in the more volatile components. *See* VAPOR PRESSURE.

The process involved in successive partial vaporizations and condensations (Fig. 2*a*) would lead to only very small amounts of the most enriched products, along with numerous streams having intermediate compositions. A logical step is to recycle each of these intermediate streams to the prior vessel in the sequence of contactors (Fig. 2*b*). Recycling of the intermediate vapors and liquids has another highly beneficial effect in that it negates the need for intermediate heaters and coolers. The resultant process (Fig. 2*b*) is known as continuous fractional distillation. It is usually carried out in a distillation column (**Fig. 3**), which is a simpler, more compact form of equipment than the cascade of vessels used in the process of recycling intermediate vapors and liquids. However, the processes are functionally equivalent.

A distillation column consists of a series of stages, also called trays or plates, on which vapor and liquid are contacted. The liquid and vapor pass through the

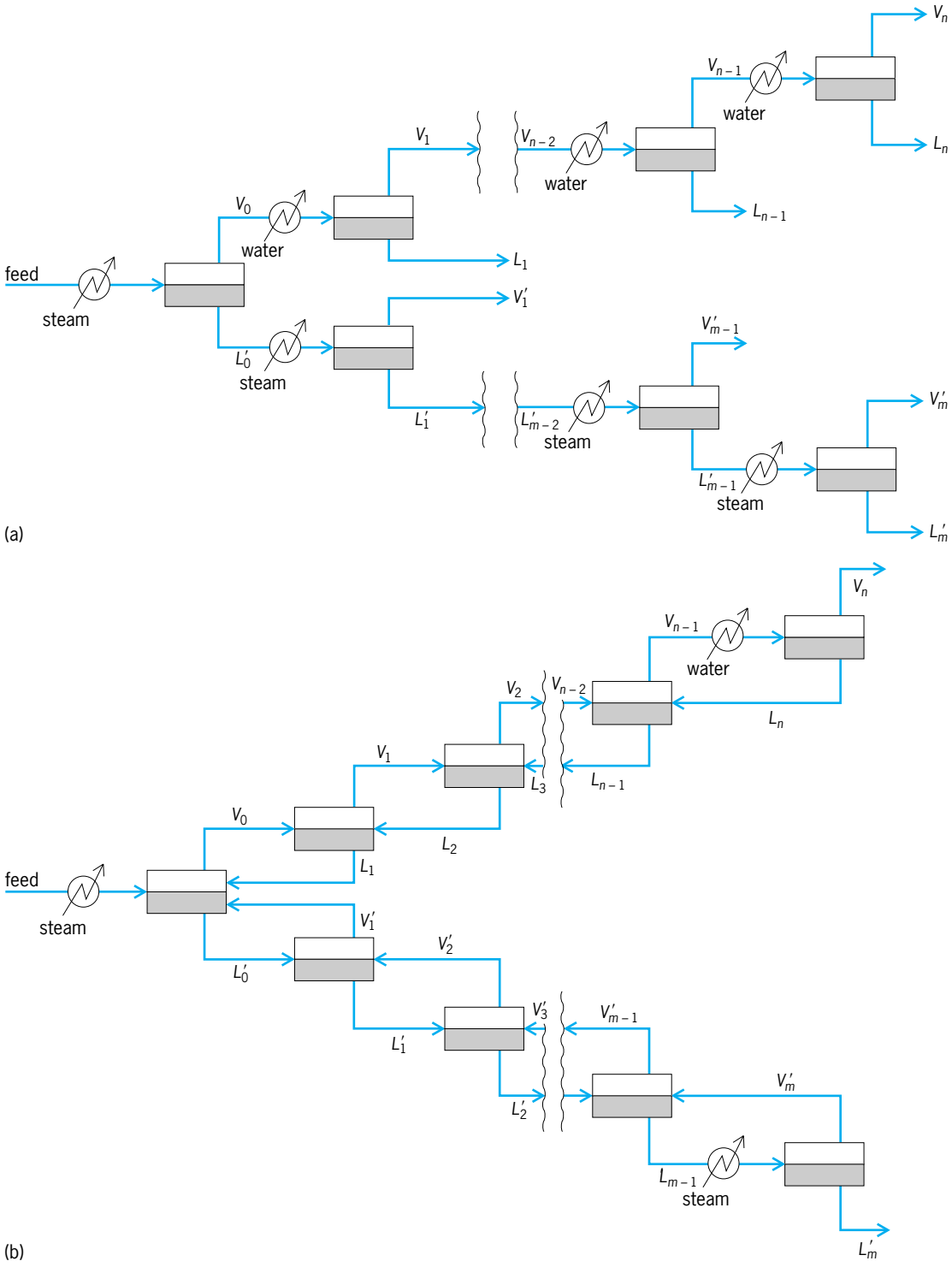

Fig. 2. Distillation by repeated vapor-liquid equilibrations. (*a*) Increasing product purities by successive partial vaporizations and condensations. (*b*) Recycling intermediate products and elimination of intermediate heating and cooling steps; equivalent to continuous fractional distillation. Symbols are explained in the text.

column in opposite directions, the liquid downward and the vapor upward, a pattern known as countercurrent flow. Liquid is held onto the trays by weirs over which the liquid exits the tray, passing into a vertical downcomer that conveys the liquid to the tray below. Vapor rises up the column through the trays successively.

Various tray designs are used. Sieve trays contain a large number of small holes, of the order of 4 mm (0.16 in.) in diameter, and the velocity of the vapor through these holes is sufficient to prevent leakage of liquid through them. Valve-cap trays contain much larger holes, of the order of 3 cm (1.2 in.), each fitted with a free-floating bonnet. Bubble-cap

trays are common in older columns and consist of a fixed bonnet, which forces the vapor out horizontally through slots. Sometimes distillation columns are constructed without discrete stages and simply have vapor and liquid flowing countercurrently through a bed of divided solid material, called packing. The form of the packing is usually chosen to give high surface area per unit bed volume, as well as a high porosity. *See* DISTILLATION COLUMN.

It is necessary for the operation of continuous fractional distillations (Figs. 2*b* and 3) that there be a cooler at the top and a heater at the bottom of the cascade of countercurrent stages. The overhead cooler is known as the condenser, and the liquid that it generates is known as reflux. The heater is termed a reboiler. The stages above the point at which the feed is supplied are known as the rectifying, or enriching, section, and they serve to remove the less volatile components from the more volatile components that become the overhead product. The stages below the feed point are known as the stripping section, since they serve to strip the more volatile components out of the less volatile components that compose the bottom product. The feed can be either vapor or liquid, or a mixture of the two. There can be multiple feeds, which may enter the column at different locations. There can also be intermediate products, or sidestreams, withdrawn.

A batch distillation can also be carried out in staged equipment. The vertical section of the batch still (Fig. 1*b*) is then equipped with trays and serves as a rectifying section, and an overhead condenser serves to generate liquid reflux, which is returned to the top of the column. Batch distillation with a rectifying section serves to remove the less volatile components to a greater extent from the overhead product; however, the absence of a stripping section means that there cannot be a similar added purification of the remaining liquid.

Vapor-liquid equilibria. The separation accomplished in a distillation relates to the difference in composition of vapor and liquid phases at equilibrium. These relationships are the subject matter of phase-equilibrium thermodynamics. *See* CHEMICAL THERMODYNAMICS; GAS; LIQUID.

Compositions of phases in equilibrium can be determined experimentally. A common device for measuring vapor-liquid equilibria is the Othmer still, in which a liquid is heated to form equilibrium vapor, which is then condensed and delivered to a distillate receiver, from which distillate liquid drains back to mix with the original liquid. Sampling and analyzing the distillate and the main liquid gives the compositions of a vapor and a liquid, respectively, in equilibrium with one another.

Powerful methods have been developed for predicting vapor-liquid equilibrium data. Vapor pressures are obtained from measurements of pure components or from any of several methods for predicting vapor pressures based on deductive reasoning. Activity coefficients are obtained from any of a variety of predictive methods; one of the more commonly used is known as the UNIFAC method.

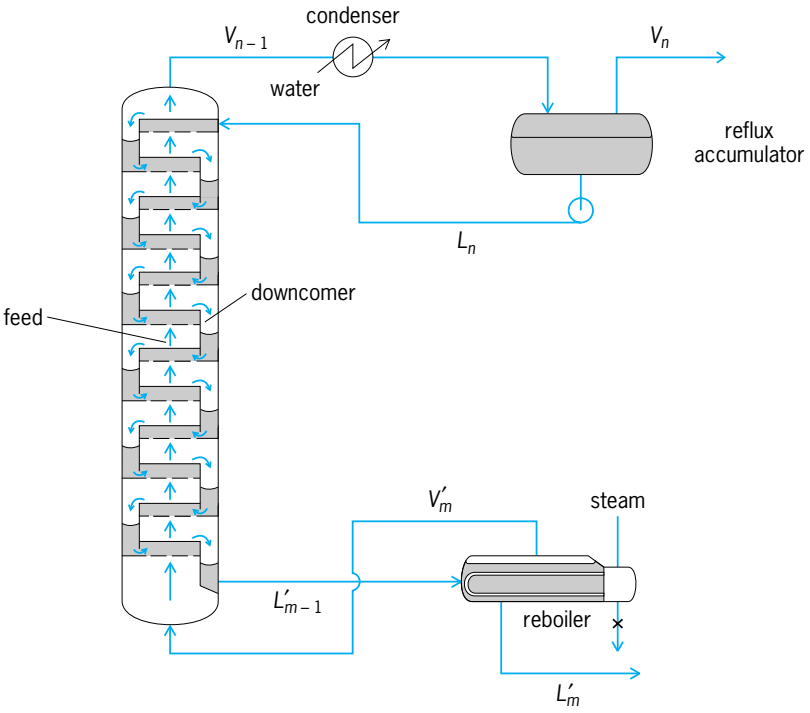

Fig. 3. Distillation column, carrying out continuous fractional distillation.

The dependence of liquid-phase activity coefficients upon liquid composition is complex; consequently vapor-liquid equilibrium relationships can have quite different characteristics. For example, the two-component benzene-toluene system (**Fig. 4**, curve A) is a relatively ideal system, for which the activity coefficients are close to 1.0 over the liquid composition range. The mole fraction of benzene in the vapor is therefore always greater than that in the

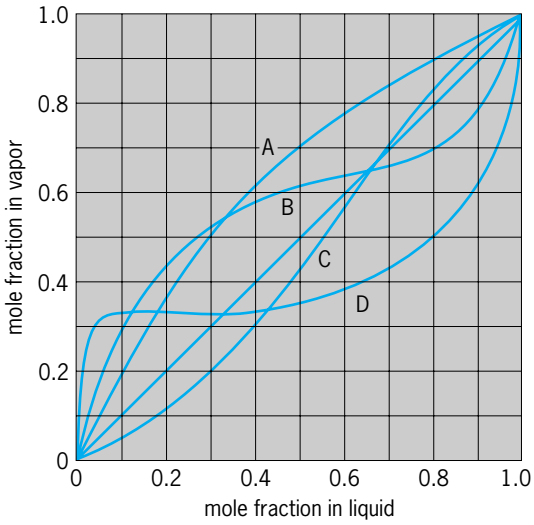

Fig. 4. Types of binary vapor-liquid equilibrium relationship; equilibrium vapor mole fraction of the more volatile (first named) component as a function of the liquid mole fraction of the more volatile component. The pressure is constant at 1 atm (101 kilopascals) for all systems. Curve A, benzene-toluene; B, acetone-carbon disulfide; C, acetone-chloroform; D, isobutyl alcohol-water.

liquid, and the relative volatility has a nearly constant value of 2.4. Such a system can be separated readily into products of high purity by continuous fractional distillation.

In the acetone-carbon disulfide system (Fig. 4, curve B), there are positive deviations from ideality, meaning that the activity coefficients are greater than 1.0 and increase for either component as that component becomes more dilute in the mixture. The result is that the relative volatility is greatest at low concentrations of the more volatile component (acetone) and decreases at higher concentrations. In this case, the curve crosses the line at which the vapor mole fraction (y_A) equals the liquid mole fraction (x_A) at 0.67, and the relative volatility (α_{AB}) becomes less than 1.0 at higher values of x_A and y_A. There is therefore a minimum-boiling azeotrope with a composition $x_A = 0.67$. Continuous fractional distillation cannot produce two products of high purity for such a system. A composition approaching that of the azeotrope is created as distillate, and depending upon the feed composition, either relatively pure A or relatively pure B is produced as bottoms. *See* AZEOTROPIC MIXTURE.

Because of hydrogen bonding between the two components, the acetone-chloroform system (Fig. 4, curve C) exhibits negative deviations from ideality, meaning that the activity coefficients are less than 1.0. Consequently, the relative volatility is less than 1.0 at low x_A and increases to greater than 1.0 at high x_A. A maximum-boiling azeotrope occurs at $x_A = 0.66$. Continuous fractional distillation cannot cross the azeotropic composition. It produces either relatively pure A or relatively pure B as distillate and a composition approaching that of the azeotrope as bottoms.

The isobutanol-water system (Fig. 4, curve D) exhibits sufficiently extreme positive deviations from ideality so that two immiscible liquid phases form over much of the composition range. Over the range of x_A where two immiscible liquids are formed, the equilibrium y_A is constant, and the curve is flat. The y_A-x_A equilibrium relationship crosses the $y_A = x_A$ diagonal within this region. The system, therefore, forms what is known as a heterogeneous azeotrope at $y_A = 0.34$, where y_A is equal to the hypothetical x_A that would obtain if the two immiscible liquids were to have formed a single liquid phase.

Distillation pressure. The pressure under which a distillation is performed is a matter of choice, although operation at pressures more removed from atmospheric becomes more costly because of the needs for pressure vessels at high pressure and for large column diameters and means of producing vacuum at subambient pressures. Often the column pressure is dictated by the pressure at which the feed is available or at which the products are to be utilized. Altering the pressure of a distillation can serve to alter the vapor-liquid equilibrium relationship, a feature that can be used to advantage. For example, the minimum-boiling azeotrope in the ethanol-water system disappears at a low-enough column pressure.

Because of the relationship between vapor pressure and temperature, the temperatures within a distillation column are lower for lower pressures. Vacuum distillation is an effective means of maintaining lower temperatures for separations involving heat-sensitive materials.

Steam distillation is an alternative to vacuum distillation for separations of organic substances. In this process, steam is fed directly to the bottom of a column and passes upward, composing a substantial fraction of the vapor phase. The combined partial pressures of the organic substances being distilled are thereby lessened, giving the lower temperatures characteristic of a vacuum distillation.

Molecular distillation. An extreme of vacuum distillation is molecular distillation, in which a liquid is vaporized from a pool or film, with a chilled condenser surface only a small distance away from the surface of the liquid. The process is carried out at such low pressure that the distance between the liquid surface and the condenser surface is less than the mean free path of the vapor, meaning that the molecules do not collide with one another or with any other molecule on their way from the liquid to the condenser surface. Typical conditions are a pressure of 0.001 mmHg (0.13 pascal) absolute and a gap of about 0.2 cm (0.08 in.). One molecular distillation apparatus, the Hickman still, uses centrifugal force to spread a thin film of liquid across a heating surface.

Molecular distillation enables a distillative separation to occur at minimum temperature and minimum time of exposure to that temperature. It is therefore well suited for heat-sensitive materials having relatively low volatilities, typically substances in the molecular weight range of 400–1200.

Extractive and azeotropic distillation. Another way to alter relative volatilities and thereby overcome limitations stemming from azeotropes and to facilitate distillations in other ways is to add another component in an amount sufficient to alter the relative volatility of the components being separated. In

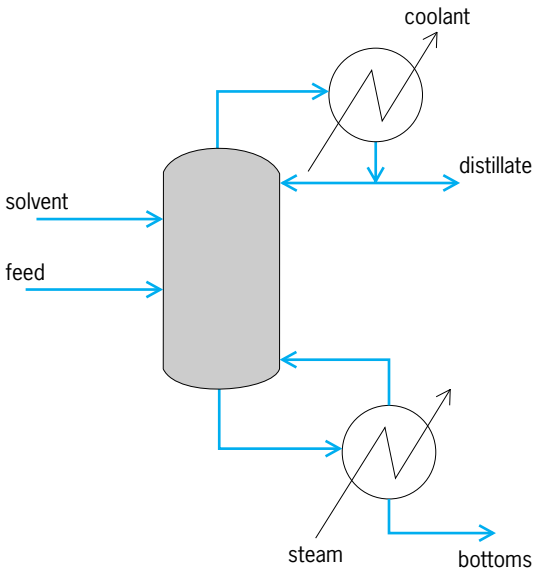

Fig. 5. Column used for extractive distillation.

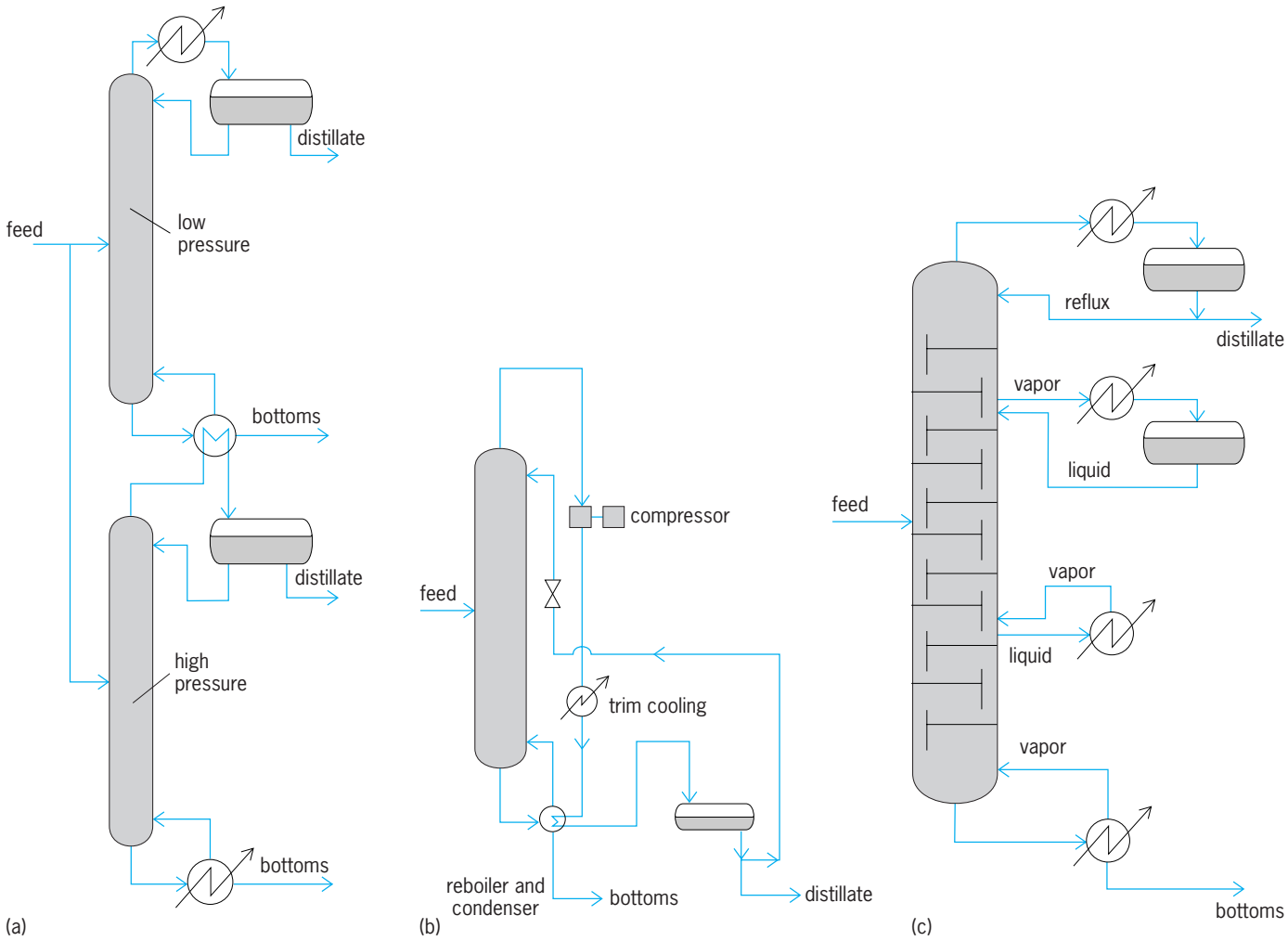

Fig. 6. Methods of reducing energy consumption in distillation. (*a*) Multieffect columns. (*b*) Vapor recompression. (*c*) Intermediate condensers and reboilers.

extractive distillation (**Fig. 5**), a solvent is added to a stage above the feed in a continuous fractional distillation. The solvent forms a substantial portion of the liquid on all stages below the addition point and alters the activity coefficients of the feed components in the liquids on those stages. If the solvent interacts with the less volatile component preferentially, the relative volatility between the feed components becomes greater, and fewer stages are needed to separate the feed components. The stages between the feed and solvent-addition points serve as the rectifying section, reducing the concentration of the less volatile component; and the stages above the solvent-addition point serve to remove solvent from the overhead distillate. A subsequent separation is needed to recover the less volatile component from the solvent.

A related process is azeotropic distillation, where an added volatile component comes overhead in the distillate. The design of the column is specified so that the added component exists at a high-enough concentration in the liquid on a number of trays to improve the relative volatility of the feed components on those stages. *See* AZEOTROPIC DISTILLATION.

Enhancing energy efficiency. Distillation consumes heat energy in the reboiler or refrigeration energy in the condenser of a column operating at subambient temperature. Often the amount of energy needed is large, and there is incentive to reduce it. Three methods have been employed to lessen energy consumption in distillation (**Fig. 6**). In one method, a distillation is carried out in two different columns, operating at different pressures (Fig. 6*a*). The difference between column pressures is such that the condenser of the high-pressure column can serve as the reboiler for the low-pressure column. Thus, the requirement for steam or other heating medium is approximately halved, although the temperature span between the high-pressure reboiler and the low-temperature condenser is approximately doubled. This approach is used in the production of oxygen and nitrogen by low-temperature distillation of air.

In a vapor-recompression system (Fig. 6*b*), the overhead vapor is compressed to a high-enough pressure that it will condense in the reboiler, thereby replacing reboiler energy with mechanical compression. This approach is most attractive for distillations of very close boiling mixtures, such as propylene

and propane, where the temperature span across the distillation is not large.

The third method uses intermediate condensers and reboilers (Fig. 6*c*). The reflux and boil-up requirements of a continuous fractional distillation are typically set by the sizes of the liquid and vapor flows needed in the vicinity of the feed stage. A less cold medium can be used in an intermediate condenser, to the extent that some of the condensation can be accomplished there rather than entirely in the overhead condenser. Similarly, a heating medium at a lower temperature can be used in an intermediate reboiler, to the extent that some of the vaporization can be accomplished there rather than entirely in the reboiler at the bottom of the column.

Computational methods. If the number of equilibrium stages, the column pressure, the descriptors of the feeds (composition, flow rate, phase condition and location), the distillate flow rate, the condenser or reboiler heat duty, and the location, phase condition, and flow rates of any sidestreams are specified, the computation of the product compositions and the compositions of liquid and vapor on all stages in a continuous fractional distillation becomes a problem of solving the so-called MESH equations. M refers to mass balances for each component for each stage; E, to vapor-liquid equilibrium relationships for each component on each stage; S, to summation of individual component flows to produce the total liquid and vapor flows leaving each stage; and H, to enthalpy balances for each stage. *See* ENTHALPY.

Historically, a number of assumptions (constant relative volatilities, constant molar liquid and vapor flows from stage to stage, and others) have often been made to simplify calculation methods for continuous fractional distillations. However, as more capable computers have increased computational power, it has become common to solve the equations in full, provided that adequate data or predictive methods are available for vapor-liquid equilibria and enthalpies. Numerous computer routines are commercially available for this purpose.

Stage efficiency. The stages, or trays, of a distillation column generally do not provide equilibrium between the vapor and liquid exiting the stage. There are two primary reasons. First, the residence times of the vapor and liquid and the intensity of contacting are usually not sufficient for equilibrium to be achieved. This effect reduces the separation as compared with simple equilibrium. Second, the liquid typically flows across the tray, and the vapor equilibrates against liquids of all compositions from inlet to outlet. This effect enhances the separation as compared with simple equilibration. Prediction of stage efficiencies is complex and is typically the area of most uncertainty in the design of distillation columns. To a large extent, it is still based upon experience, although research is continually providing insight into the flow, mixing, and mass-transfer phenomena involved. *See* FIRED HEATER. C. Judson King

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Distillation column

An apparatus used widely for countercurrent contacting of vapor and liquid to effect separations by distillation or absorption. In general, the apparatus consists of a cylindrical vessel with internals designed to obtain multiple contacting of ascending vapor and descending liquid, together with means for introducing or generating liquid at the top and vapor at the bottom.

In a column that can be applied to distillation (**Fig. 1**), a vapor condenser is used to produce liquid (reflux) which is returned to the top, and a liquid heater (reboiler) is used to generate vapor for introduction at the bottom. In a simple absorber, the absorption oil is the top liquid and the feed gas is the bottom vapor. In all cases, changes in composition produce heat effects and volume changes, so that there is a temperature gradient and a variation in vapor, and liquid flows from top to bottom of the column. These changes affect the internal flow rates from point to point throughout the column and must be considered in its design.

Distillation columns used in industrial plants range in diameter from a few inches to 40 ft (12 m) and in height from 10 to 200 ft (3 to 60 m). They operate at pressures as low as a few millimeters of mercury and

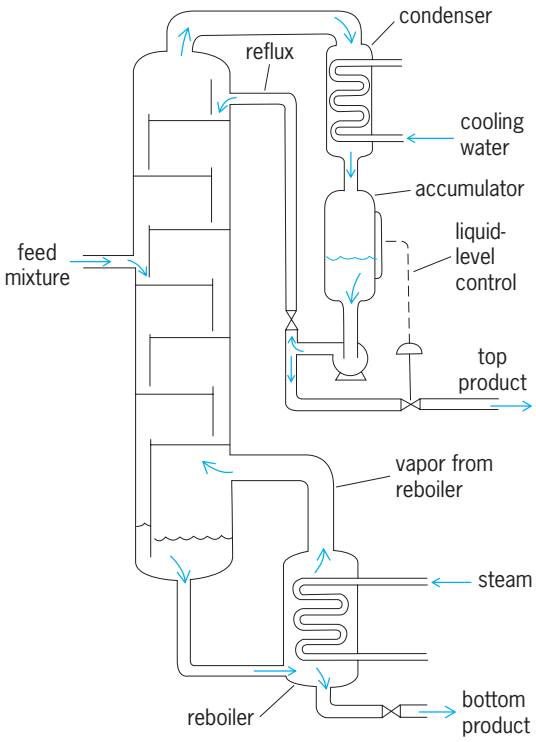

Fig. 1. Elements of a distillation column.

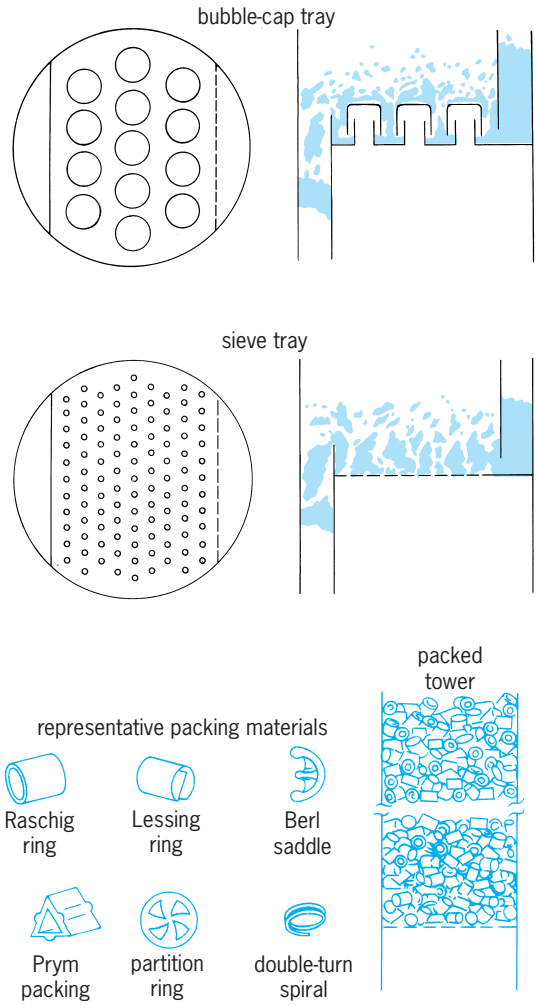

Fig. 2. Contacting devices for fractionating column

as high as 3000 lb/in.^{2} (2 megapascals) at temperatures from −300 to 700°F (−180 to 370°C). They are made of steel and other metals, of ceramics and glass, and even of such materials as bonded carbon and plastics.

A variety of internal devices have been used to obtain more efficient contacting of vapor and liquid. The most widely used devices are the bubble-cap plate, the perforated or sieve plate, and the packed column (**Fig. 2**).

Plates. The bubble-cap plate is a horizontal deck with a large number of chimneys over which circular or rectangular caps are mounted to channel and distribute the vapor through the liquid. Liquid flows by gravity downward from plate to plate through separate passages known as downcomers. The number of caps and the size of downcomers are designed to fit the expected internal flow rates of liquid and vapor.

The bubble-cap plate is the most common type of contacting device used in distillation columns. It can handle a wide range of vapor and liquid flows, is not obstructed by small amounts of rust and other solids, and is manufactured by a large number of suppliers.

The perforated or sieve plate is a horizontal deck with a multiplicity of round holes or rectangular slots for distribution of vapor through the liquid. The sieve plate can be designed with downcomers similar to those used for bubble-cap trays, or it can be made without downcomers so that both liquid and vapor flow through the perforations in the deck. The simple sieve tray has the disadvantage of a restricted operating range, so that operation at vapor flows much less than the design rate results in poor contacting.

A number of variations and improvements have been made in the sieve tray to improve the operating range. The most successful of these is the valve tray, in which a check valve is placed over each perforation to adjust for the variation in vapor flow. The valve usually consists of a disk over each perforation which is held within a cage attached to the deck. The disk is free to move up and down within the limits imposed by the cage, and thus adjusts to the vapor flow.

Although it is older than the bubble-cap plate, the sieve tray did not obtain wide use until stainless steel sheets became readily available to overcome difficulties with corrosion and fouling. With the improvements, such as the valve tray, this type of plate may supplant the bubble-cap plate in many services.

Packings. The packed column is a bed or succession of beds made up of small solid shapes over which liquid and vapor flow in tortuous countercurrent paths. Expanded metal or woven mats are also used as packing. The packed column is used without downcomers, but in larger sizes usually has horizontal redistribution decks to collect and redistribute the liquid over the bed at successive intervals of height. The packed column is widely used in laboratories. It is often used in small industrial plants, especially where corrosion is severe and ceramic or glass materials must be chosen.

Other contacting means sometimes used in fractionating columns are the disk-and-doughnut tray for very viscous or dirty liquids, the spiral ribbon for very small columns, the spray column, and the welled-wall column. *See* GAS ABSORPTION OPERATIONS; DISTILLATION.

Mott Souders

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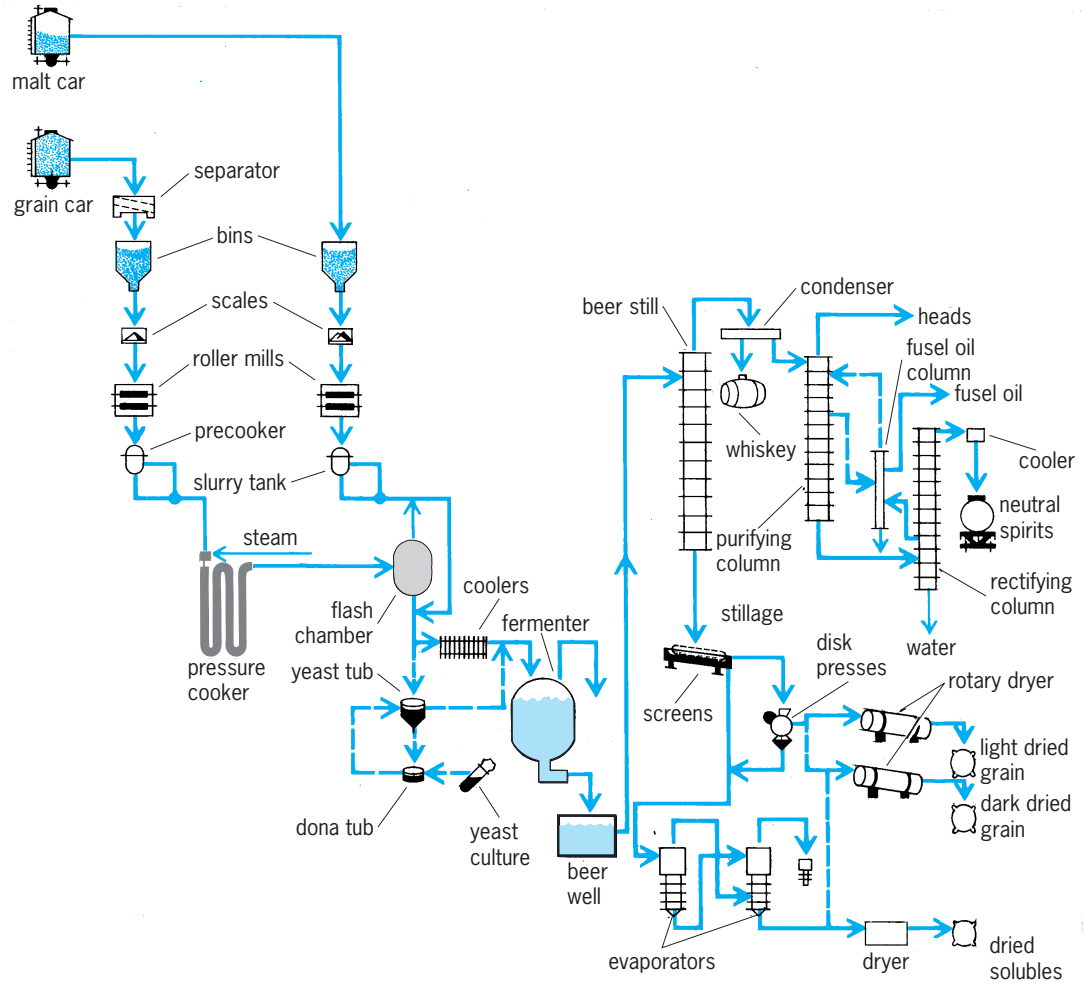
Distilled spirits

Potable alcoholic beverages, each obtained by distilling an alcohol-containing liquid and further treating the distillate to obtain a beverage of specific character.

Classification. The various distilled spirits are classified according to (1) the raw material which has been fermented and subsequently distilled, such as grain, molasses, and fruit; and (2) the further treatment given the distillate to add specific flavors



Flow diagram of whiskey production. (*J. E. Seagram and Sons*)

and aroma. These are referred to as compounded or flavored spirits. Examples of beverages made from grain are whiskeys, vodka, and gin. Some types of vodka and gin are made from potatoes. Molasses is used for the manufacture of rum, and sugarcane juice for the Brazilian cachaça or pinga. Fruits are used for the preparation of brandies, and agave juice for Mexican tequila. The so-called compounded or flavored spirits are made from distilled spirits by the addition of sugar, essential oils, color, herbs, bitters, or other ingredients. Examples are English and Dutch gins, various liqueurs or cordials, absinthe, and aquavit.

Fermentation. When starchy materials are to be fermented, it is necessary to convert the starch to fermentable sugars first. This is done by the addition of amylolytic or diastatic enzymes from malt or certain fungi. Once the starch has been converted to sugar, mostly maltose and glucose, it can be fermented by yeast to alcohol and carbon dioxide. When wine is distilled the fermentation is carried out by various strains of the wine yeast, *Saccharomyces cerevisiae* or its variety *ellipsoideus*. With other sugar-containing substrates, special distillery yeasts may be employed, usually also strains of *S. cerevisiae*, or a natural fermentation by uncontrolled species of yeast is used. *See* FERMENTATION; WINE.

Distillation process. This is done in stills of various design. In a simple pot still very little fractionation, or rectification, is accomplished. When a still is equipped with a rectifying column, fractions of much higher purity can be collected (see **illus.**). The batch-type distillation is becoming replaced more and more by continuous stills, especially in the production of grain alcohol or neutral spirits. Normally, three types of fractions are separated. A fraction with a low boiling point, the "heads," is rich in aldehydes and esters. The main fraction is ethyl alcohol and a higher-boiling fraction is rich in so-called fusel oils or higher alcohols. The latter consist of D-amyl- and isoamylalcohol, butyl- and isobutylalcohol, propanol, and others. Isoamylalcohol and amylalcohol are the principal ingredients of fusel oil. The strength of the ethyl alcohol is expressed by the term "proof." In the United States 100 proof spirit is a spirit containing 50% alcohol by volume at a temperature of 60°F (16°C).

Types. The various types of distilled spirits are discussed below.

Brandies. Brandy is a spirit obtained from the distillation of wine or a fermented fruit juice, usually after aging of the wine in wooden casks. Cognac is a brandy distilled from wines made of grapes grown within the legal limits of Charente and Charente Inférieure Departments, the Cognac region of France. Armagnac is brandy made in the Department of Gers, southeast of Bordeaux. Both types of brandy are aged for many years in oak before bottling. Before bottling, various lots are blended, the alcohol content is adjusted to 42–44% and if necessary, coloring matter such as caramel is added. Brandies distilled from grape pomace of the wine press are called *eau de vie de marc* in France and *grappa* in Italy. Spanish brandies are usually distilled from sherry wines and have a distinctive flavor, quite different from cognac or armagnac. American brandies are primarily products of California and have a flavor different from the European brandies. Whereas in Europe pot stills are most common, continuous stills are preferred in California. Apple brandy is called applejack in the United States and is called calvados in France. It is distilled from completely fermented apple juice and aged in oak barrels for 5–10 years; it has a distinct apple flavor. Other fruits from which brandy is made include black, wild cherries (kirsch or kirschwasser), plums (slivovitz from Hungary or Rumania, and quetsch or mirabelle from France), blackberries, and apricots. When stone fruits are used, some of the stones are broken or crushed and a small amount of the oil is distilled over with the spirit, giving the brandy a more or less pronounced bitter almond flavor.

Whiskeys. These distilled spirits are made by distilling fermented grain mashes and aging the distillate in wood, usually oak. Examples are Scotch whiskey, Irish whiskey, Canadian whiskey, rye whiskey and bourbon whiskey.

Scotch whiskey is made primarily from barley. The barley is converted to malt by allowing it to sprout after which it is dried in a kiln over a peat fire. The malt absorbs some of the smoke aroma which is carried over later with the spirit distilled from it. The drying and roasting of the malt is to a large extent responsible for the flavor of the whiskey. After the malt is made into a mash, it is fermented, distilled, and aged in oak casks. These are often casks in which sherry has been shipped. A Scotch blended whiskey may contain a certain percentage of grain whiskey besides malt whiskey.

Irish whiskey is made from malt, unmalted barley, and other grains such as wheat, rye, and oats. The malt is not smoke-cured as in Scotland, and as a result the flavor of the resulting whiskey is different from that of Scotch whiskey.

Many types of whiskey are made in the United States. Depending on the principal source of grain, whiskeys are classified as bourbon (corn or grain), rye whiskey (rye), and others. Blended whiskeys are differentiated from straight whiskeys by a certain content of neutral spirits. American whiskeys are aged for a number of years in new charred white-oak barrels. In all types of whiskeys discussed above, the type and quality of the water used in the plant is an important factor in the quality of the finished product.

Gins. These consists essentially of a pure grade of alcohol with the addition of an extract of the juniper berry as the chief flavoring agent. The flavor is imparted by distillation of herbs (distilled gin) or by adding essential oils (a compounded gin). There are two principal types: English, or London dry, and Dutch (jenever).

English gin is made from pure neutral spirits, which are redistilled in the presence of juniper berries and small amounts of other ingredients such as coriander seed, cardamon seed, orange peel, anise seed, cassia bark, fennel, and others. English gin is not aged.

Dutch gin is made of malt. The flavoring ingredients are mixed directly in the mash, after which it is distilled at a rather low proof. Dutch gins are heavier in body and contain more fusel alcohols and other volatile compounds (congenerics) besides ethyl alcohol than does English gin. Some types of Dutch gin are aged. Dutch gin has a pronounced flavor of its own and is consumed straight, whereas English gin is usually used as an ingredient in cocktails.

Rum. The alcoholic distillate from fermented sugar cane juice or molasses is known as rum. Cuban rum is light-bodied and light-colored. The middle fraction of the distillation, known as aguardiente, is used for making rum. The aguardiente is aged in uncharred oak barrels; it is then decolorized, filtered, and supplied with some caramel to give it the proper color. Occasionally some fruit aroma is added. It is then aged for several more years. Jamaica rum is heavier-bodied and contains more congenerics (fusel alcohol, esters, and aldehydes) than Cuban rum. Arak is a rum that comes from the island of Java. It is a dry, highly aromatic rum. A natural fermentation of the molasses by various species of yeast also contributes to the flavor of this drink.

Vodka. A product originally produced in Russia but now popular in many countries, vodka is usually made from wheat. It is highly rectified during distillation and thus is a very pure neutral spirit without a pronounced taste. It is not aged.

Cachaça or pinga. This is Brazilian spirit made by distilling naturally fermented sugarcane juice in pot stills. It is high in congenerics and sold at various degrees of proof. It is usually not aged.

Pulque and tequila. Pulque and tequila are of Mexican origin. A sweet sap is obtained from agave (century plant and American aloe) by removing the growing bud from about 3-year-old plants. The sap (aguamiel or honey water) is collected and fermented by natural fermentation and is then called pulque. Distilled pulque is tequila.

Liqueurs. Liqueurs or cordials are alcoholic beverages prepared by combining a spirit, usually brandy, with certain flavorings and sugar. In fruit liqueurs, the color and flavor are obtained by an infusion process using the specified fruit and spirit. Sugar is added after the extraction is complete. Plant liqueurs

are made by maceration of plant leaves, seeds, or roots with spirit and then distilling the product. Sugar and coloring matter are added after distillation. A large variety of different liqueurs are marketed. *See* YEAST.

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Distortion (electronic circuits)

The behavior of an electrical device or communications system whose output is not identical in form to the input signal. In a distortionless communications system, freedom from distortion implies that the output must be proportional to a delayed version of the input, requiring a constant-amplitude response and a phase characteristic that is a linear function of frequency. *See* RESPONSE.

In practice, all electrical systems will produce some degree of distortion. The art of design is to see that such distortion is maintained within acceptable bounds, while the signal is otherwise modified in the desired fashion. In analog systems operating in the audio band, major emphasis has been placed on satisfying the constant-amplitude characteristic since the human ear is insensitive to changes in phase. However, in digital systems, phase distortion can lead to unacceptable timing problems. In general, distortion can be grouped into four forms: amplitude (nonlinear), frequency, phase, and cross modulation.

Amplitude distortion. All electronic systems are inherently nonlinear unless the input signal is maintained at an incrementally small level. Once the signal level is increased, the effects of device nonlinearities become apparent as distorted output waveforms. Such distortion reduces the output voltage capability of operational amplifiers and limits the power available from power amplifiers. *See* AMPLIFIER; OPERATIONAL AMPLIFIER.

The load-line characteristic for a simple transistor amplifier is extremely helpful in predicting the output waveform for resistive loads if the input is sinusoidal in form. Excursions of the input signal about the quiescent operating point of the device are readily translated to output excursions. In this way the maximum input signal level can be determined for the avoidance of undesirable output distortion.

When the load is not resistive, however, or when more precise information is required, power series representation of the device is utilized. Thus, for example, the relationship between a transistor collector current i_c and its base current i_b may be expressed by Eq. (1). Thus, if the first two terms are used, and

$$i_c = G_1 i_b + G_2 i_b^2 + \cdots + G_n i_b^n + \cdots \qquad (1)$$

the input is sinusoidal, the output contains both the fundamental and a second harmonic. The degree of distortion can then be determined by comparing the relative amplitudes of the harmonic and fundamental terms. *See* HARMONIC (PERIODIC PHENOMENA); TRANSISTOR.

Thus, if the fundamental term has amplitude A_1 and the ith harmonic has amplitude A_i, harmonic distortion may be defined by Eq. (2). A measure of

$$D_i = \frac{A_i}{A_1} \qquad i \neq 1 \qquad (2)$$

the total harmonic distortion (THD) is then given by Eq. (3). This quantity is most commonly associated

$$D = \sqrt{D_2^2 + D_3^2 + \cdots} \qquad (3)$$

with an output voltage signal.

A common form of distortion in operational amplifier systems results from the slew-rate limitation of the amplifier. To conserve quiescent current, many operational amplifiers are designed to have a class B output stage capable of driving as much as 50 times its quiescent current. This puts demands on the intermediate drive stage, which is required to slew rapidly through a one- or two-diode voltage drop during transition from the push to pull mode, and vice versa. For slew rates beyond its capacity (usually arising because of an inability to drive the output junction capacitance), the output exhibits slew-rate limiting. This takes the form of straight-line segments in sinusoidal outputs, and can result in triangularization of the waveform if severe slew-rate limiting is encountered. For square waves, slew-rate limiting begins with finite transition rates between low and high states, with a triangular waveform occurring in extreme cases. For commercially available operational amplifiers, slew rates are typically in the range 0.5–20 V/microsecond and are a function of amplifier design geometry and fabrication method.

Frequency distortion. No practical device or system is capable of providing constant gain over an infinite frequency band. Hence, any nonsinusoidal input signal will encounter distortion since its various sinusoidal components will undergo unequal degrees of amplification.

The effects of such distortion can be minimized by designing transmission systems with a limited region of constant gain. Thus, in high-fidelity systems, the amplifier response is made wide enough to capture all the harmonic components to which the human ear is sensitive. In a communications system, the channel filters are designed to have a very sharp transition between regions of signal capture and signal rejection so that there is no interference between adjacent carrier bands.

Phase distortion. Since the time of propagation through a system varies with frequency, the output may differ in form from the input signal, even though the same frequency components exist. This can easily be demonstrated by noting the difference between the addition of two in-phase sine waves and

two whose phase relationship differs by several degrees.

In digital systems, such changes can be significant enough to cause timing problems. Hence, the phase–frequency response must be made linear to obtain distortionless transmission. In practice, this can be achieved by the introduction of delay equalizers whose characteristics are designed to compensate for phase distortion in other parts of the system without affecting the magnitude characteristic of the total system. *See* EQUALIZER.

Cross modulation. Sometimes referred to as intermodulation, this occurs because of the nonlinear nature of device characteristics. Thus, if two or more sinusoidal inputs are applied to a transistor, the output will contain not only the fundamentals but also signal harmonics, sums and differences of harmonics, and various sum or difference components of fundamental and harmonic components. While these effects are generally undesirable, they may be utilized to advantage in amplitude modulation and diode detection (demodulation). *See* AMPLITUDE-MODULATION DETECTOR; AMPLITUDE MODULATOR.

Reduction of distortion by negative feedback. Amplitude distortion may be reduced in amplifier stages by the application of negative feedback. If D is the distortion voltage output with feedback, and d is the distortion voltage without feedback, it can easily be shown that D is less than d by a factor of $(1 + A\beta)$, where A is the amplifier open-loop gain and β is the feedback factor. *See* FEEDBACK CIRCUIT.

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Distributed systems (computers)

A distributed system consists of a collection of autonomous computers linked by a computer network and equipped with distributed system software. This software enables computers to coordinate their activities and to share the resources of the system hardware, software, and data. Users of a distributed system should perceive a single, integrated computing facility even though it may be implemented by many computers in different locations. This is in contrast to a network, where the user is aware that there are several machines whose locations, storage replications, load balancing, and functionality are not transparent. Benefits of distributed systems include bridging geographic distances, improving performance and availability, maintaining autonomy, reducing cost, and allowing for interaction. *See* LOCAL-AREA NETWORKS; WIDE-AREA NETWORKS.

Key Characteristics

The characteristics that are primarily responsible for the usefulness of distributed systems are resource sharing, openness, concurrency, scalability, fault tolerance, and transparency. System and application software must be carefully designed to ensure that these characteristics are attained.

Resource sharing. Resources include hardware such as disks and printers, and software such as files, databases, and compilers. Sharing reduces cost and is necessary to support groups of users working cooperatively. Resources are physically encapsulated within one or more of the computers in a distributed system and can be accessed from other computers only by communications. Each set of resources of a particular type must be managed by a program called a resource manager which offers a communication interface to enable accessing, manipulating, and updating the resource. The resource manager is also responsible for providing a naming scheme for each type of resource, for enabling individual resources to be accessed from any location, for mapping of resource names to communication addresses, and for synchronizing concurrent accesses to ensure consistency. *See* COMPUTER PERIPHERAL DEVICES; SOFTWARE.

Openness. This characteristic determines whether the distributed system can be extended in various ways without disruption or duplication of existing services. Openness is achieved by specifying and documenting the key software interfaces of a system and publishing them, thus making them available to software developers. UNIX is an example of a relatively open system design. The development of a standard interprocess communication (IPC) mechanism in BSD (Berkeley Software Distribution) UNIX and other operating systems and of standard communication protocols increased the scope for achieving openness; it became possible for processes on any computer and any operating system to make their resources accessible to other processes on other computers in the network.

Concurrency and parallelism. These characteristics arise in distributed systems due to the separate activities of users, the independence of resources, or the location of processes in separate computers. If there are M computers in a distributed system with one central processing unit each, then up to M processes can run in parallel (execute simultaneously). If several processes access the same resource concurrently, their actions must be synchronized to ensure that they do not conflict. *See* CONCURRENT PROCESSING.

Scalability. Distributed systems must operate effectively and efficiently at many different scales: from a workstation and a file server, to a local-area network with several hundred workstations and many file servers, to several local-area networks connected to form an internetwork. System and application software should not need to change when the scale of the system increases. Also, the work involved in processing any single request to access a shared resource should be independent of the size of the network. The design philosophy of distributed systems is that no single resource, hardware or software, is assumed to be in restricted supply; as demand for a resource

grows, it should be possible to extend the system to meet it.

Fault tolerance. When faults occur in hardware or software, programs may produce incorrect results or they may stop before they have completed the intended computation. In a distributed system, hardware that is essential to the continued operation of critical applications can be replicated. This hardware is exploited for noncritical activities when no faults are present. Software recovery involves the design of software so that permanent data (files and other material stored on permanent storage) can be recovered when a fault is detected. *See* FAULT-TOLERANT SYSTEMS.

Transparency. This is the concealment from the user and the application programmer of the separation of components in a distributed system so that the system is perceived as a whole rather than a collection of independent components. Some examples of transparency include access transparency, which enables local and remote information objects to be accessed using identical operations; and location transparency, which enables information objects to be accessed without knowledge of their location. Access and location transparencies are referred to together as network transparency.

Design Issues

Several important design issues arise specifically from the distributed nature of systems, including naming, communication, software structure, workload allocation, and consistency maintenance. There are also design issues unrelated to a system's distribution, including interface design and software engineering.

Naming. Distributed systems are based on the sharing of resources and on the transparency of resource distribution. Names assigned to resources must have global meanings that are independent of location, and must be supported by a name interpretation system that can translate names to enable programs to access the resources. The design issue here is to design a naming scheme that will scale and translate names efficiently to meet appropriate performance goals. A process that requires access to a resource that it does not manage must possess a name or an identifier for it; in this context, "name" refers to names that can be interpreted by users or programs, and "identifier" refers to names that are interpreted or used only by programs. An appropriate name space must be chosen for each resource type. A hierarchical name space, for example, "/etc/passwd," is potentially infinite and enables a system to grow, while a flat name space is usually finite, since its size is determined by the maximum permissible length of names.

To resolve names, a name service is needed. The name service accepts requests to translate names in one name space to names in another name space. It manages a database, which provides a view of the names existing in the distributed system and their translations in other name services. A resource name is resolved when it is translated into a form in which it can be used to invoke an action on the resource to which the name refers. Names are resolved relative to some context. For example, the name John Smith does not have much meaning unless some contextual information is provided, such as the name of the organization in which he works, or the street where he lives. In file systems, each directory represents a context. For example, "/etc/passwd" is a name with two parts. The first part, "etc," is resolved relative to the context "/" or root. The name "/oldetc/passwd" can have a different meaning because its second part is resolved in a different context.

In a distributed system, a resolved resource name is a communication identifier. For example, a communication identifier on the Internet consists of an IP address (for example, 192.135.231.4), which identifies the host, and a port number, which identifies a particular communication. The communication identifier must be acceptable to the communications subsystem and is used to transmit a request to a resource manager. The communications subsystem may have to perform further translations to produce network addresses and routing information acceptable to lower-level software layers. *See* INTERNET.

Communication. Since the components of a distributed system are both logically and physically separate, they must communicate in order to interact. Distributed systems and applications are composed of separate software components that interact to perform tasks. It will be assumed that components that require or provide access to resources are implemented as processes; that is, each component is active and has its own address space. Communication between a pair of processes involves the transfer of data from the environment of the sending process to the environment of the receiving process. In some communication operations, the synchronization of the receiving process with the sending process is required, so that the sending or receiving process is prevented from continuing until the other process makes an action that frees it.

The sending and receiving processes together perform message-passing actions. This involves the transmission of a set of data values (message) through a specified communication mechanism (port or channel) and the acceptance of the message by the receiving process. The mechanism may be synchronous, wherein the sender waits after transmitting a message until the receiver has performed a receive operation; or asynchronous, wherein the message is placed in a queue of messages waiting for the receiver to accept them and the sender can proceed immediately. In distributed systems, the main patterns of communication are client-server and group multicast.

Client-server. This form of communication is oriented toward service provision. An exchange consists of (1) transmission of a request from client to server; (2) execution of the request by the server; (3) transmission of the reply to the client. The server must become aware of the request sent as soon as it arrives, and the client must wait until the reply has been received. In its lifetime, a server serves many

clients and it need have no prior knowledge of them. Each request contains a communication identifier that is used to transmit the reply to the client. In an open system, clients cannot be initialized to hold the communication identifiers of all the servers they want to access. Typically, when a server starts up, it registers itself with a name service, stating its network address and a well-known name for the service it provides. Clients obtain the server's address by interrogating the name service using the well-known name. *See* CLIENT-SERVER SYSTEM.

Group multicast. The target of this type of message is a group of processes. To a single group send operation, there corresponds a receive performed by each member of the group. Tasks performed by group multicast include:

1. Locating an object. A client multicasts a message containing the name of a file directory to a group of file servers. Only the one that holds the relevant directory replies.

2. Fault tolerance. A client multicasts a request to a group of servers, all of which process the request identically. As long as at least one of the servers does not fail, the service will be provided.

3. Multiple update. An event such as time can be multicast to a group.

Software structure. Each category of software corresponds to a different level of abstraction of the distributed system, and each offers a different degree of transparency to the user. A programmer may write in a distributed programming language, and use the underlying services provided. Besides applications, the main categories of software in distributed systems are the following.

Operating system kernel. This encompasses services that provide basic resource management, mainly memory allocation and protection, process creation, processor scheduling, interprocess communication, and peripheral device handling. It presents a set of universal problem-oriented programming abstractions, such as processes, to applications, shielding them from the underlying details of the processor and other hardware.

Open services. These bring the programming facilities of a distributed system up to the level required for application programming. New services can be developed and installed as required, enabling distributed systems to be adapted to the needs of many different types of applications. Services range from a basic file service to an interactive multimedia conferencing service.

Distributed programming support. Such software provides run-time support for language facilities, such as remote procedure calls (RPC) and group multicast communication, that allow programs written in conventional languages to work together in a distributed system environment.

Workload allocation. This addresses how work is allocated among resources in a distributed system.

Workstation-server model. The advantage of this model is putting the processor cycles near the user, especially for highly interactive applications. Processor performance and memory capacity of the workstation determine the size of the largest task that can be performed on behalf of the user; a single user with a large computing task is not able to obtain additional resources.

Processor pool model. This is a modification of the workstation-server model to include processors that can be allocated to users' tasks dynamically. A processor pool consists of a collection of inexpensive computers, each consisting of a processor, memory, and network interface. Each processor in the pool has an independent network connection. Processors do not have to be homogeneous; a system might include chip sets from different manufacturers in the pool, to broaden the range of software that can be run. Processors are allocated to processes for their lifetimes.

Use of idle workstations. This is another variation on the workstation-server model. At any one time, and especially overnight, a significant proportion of workstations on a network may be unused or be used for lightweight activities such as document editing. These stations have spare processing capacity, and can be used to run jobs for users who are logged on at other stations and do not have sufficient capacity at their machines.

Shared-memory multiprocessors. These contain several independent processors, each able to execute a separate program. Each processor interfaces to the shared memory via cache memory. A large number of central processing units can be deployed. This processing capacity is particularly suited to applications in areas such as virtual reality and multimedia. *See* MULTIMEDIA TECHNOLOGY; MULTIPROCESSING; VIRTUAL REALITY.

Consistency maintenance. This encompasses issues of consistency with regard to updates, replication, caches, failures, and user interfaces.

Update consistency. This issue arises when several processes access and update data concurrently. The activity of changing a related set of data values cannot be performed instantaneously. The updates should appear to be atomic; that is, a related set of changes made by a given process should appear to all other processes as though it was instantaneous. This is a significant issue in distributed systems because many users are likely to access shared data. Moreover, the operation of the system itself depends on the consistency of file directories managed by file services, naming databases, and so forth. *See* DATABASE MANAGEMENT SYSTEM.

Replication consistency. If data derived from a single source have been copied to several computers and subsequently modified at one or more of them, the possibility of inconsistencies arises between the values of data items at different computers. The consistency of replicas is generally ensured by periodic updates.

Cache consistency. This refers to the problem that arises when data values that have been cached by one client are updated by another. For example, the same blocks of a file stored in a file server are requested and cached at several workstations. When a client updates a block at one of the workstations,

the version held in the file server and those held in cache at other workstations are said to be stale. The client may send the updated block to the file server responsible for the file, but the other workstations will retain the out-of-date version unless the server notifies them.

Failure consistency. When one component in a distributed system fails, the others will normally continue to run. But if they depend on the continued operation of the failed computer or program, they may fail later. Each of the cooperating components involved in a failed computation may have reached a different point in its program. To ensure that the permanent data stored by all of the components remains consistent, recovery procedures are needed to cause the data to be "rolled back" to a known state after a failure.

User interface consistency. Whenever a user performs an input action such as a key depression or a mouse click in an interactive program, the screen becomes temporarily inconsistent with the user's model of the program's state. The program then processes the user's input and performs the necessary operations to update the screen. Unless the input is processed and the changes are transmitted to the screen quickly enough to give the user the impression of an instantaneous change, the user becomes aware of the screen inconsistency. In distributed systems, the transfer rates and the latency characteristics of networks have an impact on user interface consistency. *See* HUMAN-COMPUTER INTERACTION.

Object-Oriented Distributed Systems

The object-oriented model for a distributed system is based on the model supported by object-oriented programming languages. Distributed object systems generally provide remote method invocation (RMI) in an object-oriented programming language together with operating systems support for object sharing and persistence. Remote procedure calls, which are used in client-server communication, are replaced by remote method invocation in distributed object systems. *See* OBJECT-ORIENTED PROGRAMMING.

The state of an object consists of the values of its instance variables. In the object-oriented paradigm, the state of a program is partitioned into separate parts, each of which is associated with an object. Since object-based programs are logically partitioned, the physical distribution of objects into different processes or computers in a distributed system is a natural extension. The Object Management Group's Common Object Request Broker (CORBA) is a widely used standard for distributed object systems. Other object management systems include the Open Software Foundation's Distributed Computing Environment (DCE) and Microsoft's Distributed Common Object Manager (DCOM).

Common Object Request Broker. CORBA specifies a system that provides interoperability among objects in a heterogeneous, distributed environment in a way that is transparent to the programmer. Its design is based on the Object Management Group's object model.

This model defines common object semantics for specifying the externally visible characteristics of objects in a standard and implementation-independent way. In this model, clients request services from objects (which will also be called servers) through a well-defined interface. This interface is specified in Object Management Group Interface Definition Language (IDL). The request is an event, and it carries information including an operation, the object reference of the service provider, and actual parameters (if any). The object reference is a name that defines an object reliably.

The central component of CORBA is the object request broker (ORB). It encompasses the entire communication infrastructure necessary to identify and locate objects, handle connection management, and deliver data. In general, the object request broker is not required to be a single component; it is simply defined by its interfaces. The core is the most crucial part of the object request broker; it is responsible for communication of requests.

The basic functionality provided by the object request broker consists of passing the requests from clients to the object implementations on which they are invoked. In order to make a request, the client can communicate with the ORB core through the Interface Definition Language stub or through the dynamic invocation interface (DII). The stub represents the mapping between the language of implementation of the client and the ORB core. Thus the client can be written in any language as long as the implementation of the object request broker supports this mapping. The ORB core then transfers the request to the object implementation which receives the request as an up-call through either an Interface Definition Language (IDL) skeleton (which represents the object interface at the server side and works with the client stub) or a dynamic skeleton (a skeleton with multiple interfaces).

Architectural components. The communication between the object implementation and the ORB core is effected by the Object Adapter (OA). It handles services such as generation and interpretation of object references, method invocation, security of interactions, object and implementation activation and deactivation, mapping references corresponding to object implementations, and registration of implementations. It is expected that there will be many different special-purpose object adapters to fulfill the needs of specific systems (for example, databases).

The Object Management Group specifies four policies in which the object adapter may handle object implementation activation: shared server policy, in which multiple objects may be implemented in the same program; unshared server policy; server-per-method policy, in which a new server is started each time a request is received; and persistent server policy. Only in the persistent server policy is the object's implementation supposed to be constantly active (if it is not, a system exception results). If a request is

invoked under any other policy, the object will be activated by the OA in a policy-specific way. In order to be able to do that, the OA needs to have access to information about the object's location and operating environment. The database containing this information is called the implementation repository and is a standard component of the CORBA architecture. The information is obtained from there by the OA at the time of object activation. The implementation repository may also contain other information pertaining to the implementation of servers, such as debugging, version, and administrative information.

The interfaces to objects can be specified in two ways: either in IDL or through their addition to the interface repository, another component of the architecture. The dynamic invocation interface allows the client to specify requests to objects whose definition and interface are unknown at the client's compile time. In order to use the dynamic invocation interface, the client has to compose a request (in a way common to all object request brokers) including the object reference, the operation, and a list of parameters. These specifications—of objects and services they provide—are retrieved from the interface repository, a database that provides persistent storage of object interface definitions. The interface repository also contains information about types of parameters, certain debugging information, and so forth.

A server-side analogous to the dynamic invocation interface is the dynamic skeleton interface (DSI); with the use of this interface the operation is no longer accessed through an operation-specific skeleton, generated from an IDL specification. Instead it is reached through an interface that provides access to the operation name and parameters. (As in the dynamic invocation interface, the information can be retrieved from the interface repository). Thus the DSI is a way to deliver requests from the object request broker to an object implementation that does not have compile-time knowledge of the object it is implementing.

Interoperability. Many different ORB products are currently available; this diversity is very wholesome since it allows the vendors to gear their products toward the specific needs of their operational environment. It also creates the need for different object request brokers to interoperate. Furthermore, there are distributed and client-server systems that are not CORBA-compliant, and there is a growing need to provide interoperability between those systems and CORBA. In order to answer these needs, the Object Management Group has formulated the ORB interoperability architecture.

Implementation differences are not the only barrier that separates objects; other reasons for such a separation might include strict enforcement of security, or providing a protected testing environment for a product under development. In order to provide a fully interoperable environment, all those differences have to be taken into account. CORBA 2.0 introduces the higher-level concept of a domain, which roughly denotes a set of objects that, for some reason, be it implementation or administrative, are separated from some or all other objects. Thus, objects from different domains need some bridging mechanism (mapping between domains) in order to interact. Furthermore, this bridging mechanism should be flexible enough to accommodate both the scenarios where very little or no translation is needed (as in crossing different administrative domains within the same object request broker) and scenarios which can be less efficient but where there is the need to provide general access to object request brokers.

The interoperability approaches can be divided into mediated and immediate bridging. With mediated bridging, interacting elements of one domain are transformed at the boundary of each domain between the internal form specific to this domain and some other form mutually agreed on by the domains. This common form could be either standard (specified by the Object Management Group, for example, Internet Inter-ORB Protocol or IIOP), or a private agreement between the two parties. With immediate bridging, elements of interaction are transformed directly between the internal form of one domain and the other. The second solution has the potential to be much faster, but is the less general one; it therefore should be possible to use both. Furthermore, if the mediation is internal to one execution environment (for example, TCP/IP), it is known as a full bridge; otherwise, if the execution environment of one object request broker is different from the common protocol, each object request broker is said to be a half bridge.

Implementations of CORBA. Many implementations of CORBA are available. They vary in the degree of CORBA compliance, quality of support, portability, and availability of additional features.

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Distributed systems (control systems)

Collections of modules, each with its own specific function, interconnected to carry out integrated data acquisition and control in a critical environment.

Evolution of industrial control systems. Industrial control systems have evolved from totally analog systems through centralized digital computer-based systems to multilevel, distributed systems.

Analog and supervisory control. Originally, industrial control systems were entirely analog, with each individual process variable controlled by a single feedback

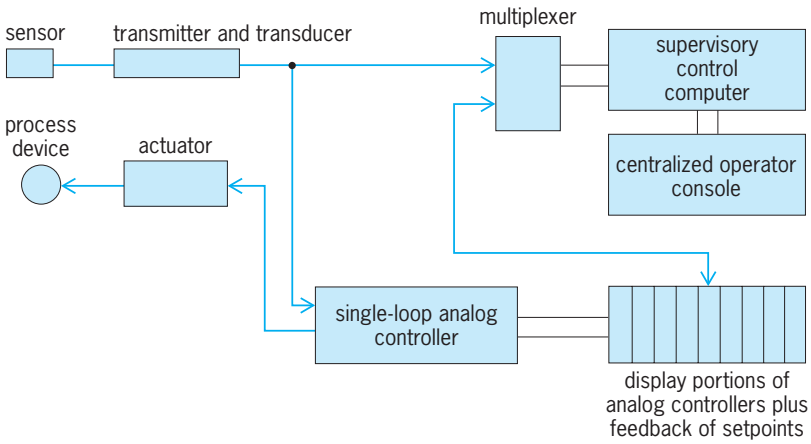

Fig. 1. Supervisory control system combining analog control with digital data acquisition and feedback of parameters to controllers.

controller. Although analog control systems were simple and reliable, they lacked integrated information displays for the process operator.

In supervisory control (**Fig. 1**), the analog portion of the system is implemented in a traditional manner (including analog display in the central operating room), but a digital computer is added which periodically scans, digitizes, and inputs process variables to the computer. The computer is used to filter the data, compute trends, generate specialized displays, plot curves, and compute unmeasurable quantities of interest such as efficiency or quality measures. Once such data are available, optimal operation of the process may be computed and implemented by using the computer to output set-point values to the analog controllers. This mode of control is called supervisory control because the computer itself is not directly involved in the dynamic feedback. *See* DIGITAL COMPUTER.

Direct digital control. Direct digital control replaces the analog control with a periodically executed equivalent digital control algorithm carried out in the central digital computer (**Fig. 2**). A direct digital control system periodically scans and digitizes process variables and calculates the change required in the manipulated variable to reduce the difference between the set point and the process variable to zero. *See* ALGORITHM.

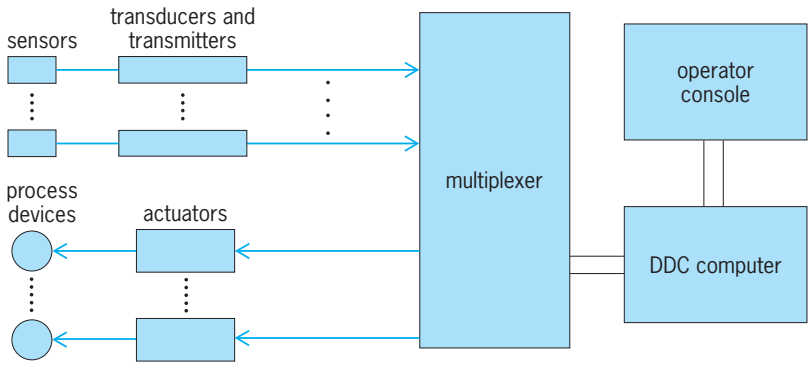

Fig. 2. Direct digital control configuration.

The advantages of direct digital control are the ease with which complex dynamic control functions can be carried out (compensation for time delay, nonlinear functions, interacting control loops) and the elimination of the cost of the analog controllers themselves. To maintain the attractive display associated with analog control systems, the display portion of the analog controller is usually provided (set-point stations). Thus operation of the process is identical to operation of digital supervisory control systems with analog controllers, except that "tuning" of the controllers (setting of gains) can be done through operator consoles.

The cost reduction which resulted from the introduction of direct digital control was offset by a number of disadvantages. The most notable of these were the decrease in reliability and the total loss of graceful degradation. Failure of a sensor or transmitter had the same effect as before, but failure of the computer itself threw the entire control system into manual operation. Hence it was necessary to provide analog controllers to back up certain critical loops which had to function even when the computer was down. As a result of this need for extra equipment, the industry tended to employ a combination of direct digital control and supervisory control, using the former only where the digital computer could perform control actions which could not be easily handled by analog controllers or where the failure of the digital computer could not jeopardize operation of the process.

Increasing demand for ever-higher levels of supervisory control, coupled with increasing installation cost, highlighted two disadvantages of centralized digital computer control of processes. First, process signals were still being transmitted from the process sensor to the central control room in analog form, meaning that separate wires had to be installed for every signal going to or from the computer. Cost of installing this cable is very high. Second, the digital computer system itself evolved into a very complex unit because of the number of devices attached to the computer and because of the variety of different programs needed to carry out the myriad control and management functions. The latter resulted in the need for an elaborate real-time operating system for the computer which could handle resources, achieve desired response time for each task, and be responsible for error detection and error recovery in a highly dynamic real-time environment. Design coding, installation, and checkout of centralized digital control systems was so costly and time-consuming that application of centralized digital control was limited.

Distributed control. Low-cost electronic hardware utilizing large-scale integrated circuits provided the technology to solve both of these problems while retaining the advantages of centralized direct digital and supervisory control. The solution, distributed control (**Fig. 3**), involved distributing control functions among hardware modules to eliminate the critical central computer. *See* INTEGRATED CIRCUITS.

Each module utilizes microprocessors in order to

carry out its function. Coupling the modules together is a digital communication system, making it feasible to geographically distribute the modules around the plant in order to minimize installation costs. This is one of the major motivations for installing such systems, since those modules which must input or output process signals may be located in the vicinity of the sensors and actuators associated with those signals and hence require minimum length of analog signal lines. Coupling modules via the digital communication system means that only one (or at most a few) communication lines need be installed around the plant. Communication links are often called "data highways" to emphasize their specialized nature and high speed. The distributed control still permits centralized collection, display, and interaction with the control system and its application through placement of operator consoles in central control rooms. *See* MICROCOMPUTER; MICROPROCESSOR.

A second motivation is the reduced design complexity of control systems, especially that of the real-time operating systems and tasks which carry out the application. The mixture of high- and low-priority tasks in one computer is replaced by tasks in separate computers or microprocessors.

A third motivation is the return of the characteristic of graceful degradation in the face of component failures. In a distributed system, loss of any module can be limited to loss of the function supplied by that module. If that function is not critical, the system merely continues without the function in a degraded performance mode. If that function is critical, it may be backed up by a duplicate module or by dynamically assigning its function to another module when the failure is detected. In any case, there is no longer the dependence on any one module for successful operation of the system, nor the need for complex and expensive backup of that module.

Multihighway distributed systems. The success of distributed control led to the demand for ever larger control systems which take the form of tightly interconnected distributed systems (**Fig. 4**). A large system by modern standards might include 10,000 or more individual sensors and as many as 30,000 computed variables. The need for timely communication of data in such large systems makes the data communication system (the data highways) very critical. Individual highways are called local highways and generally are associated with a specific unit or area of the application. The supervisory highway is usually associated with overall management of the application or plant. Its nodes are commonly operator consoles and computer interfaces, and are often in a central control room associated with the overall plant.

Module functions. A distributed control system (Fig. 3) is made up of tightly interconnected modules, each with its own specific function.

Data acquisition. A data acquisition module accepts some number of analog input signals, scans and digitizes them at arbitrary rates, performs alarming and smoothing calculations, and transmits the reduced digital data to other modules in the system for direct or supervisory control purposes. Binary (on/off) signals may also be acquired.

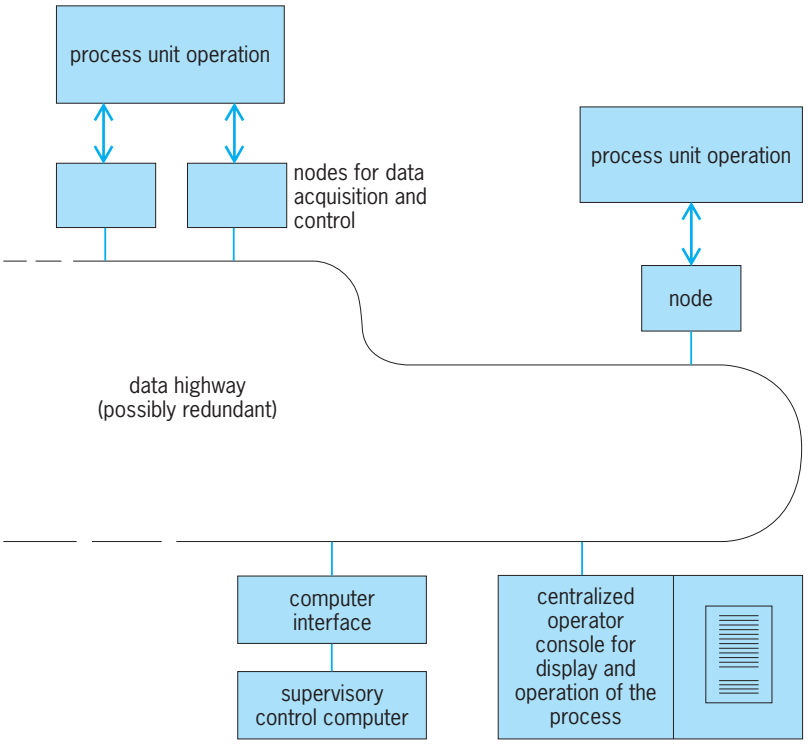

Fig. 3. Distributed data acquisition and control system with geographical distribution of nodes.

The timing of acquisition of a signal's value is often very critical to the application. For example, the signal might have to be acquired periodically with very small variation of the periods between signal value acquisitions, as in direct digital control.

Direct digital controllers. A direct digital controller performs exactly the same functions on one or a few loops as those carried out in a direct digital control central computer. These include digitizing, smoothing, alarming, control calculations, output of actuator signals, handling of transitions from manual to feedback control, and error recovery. Typical digital

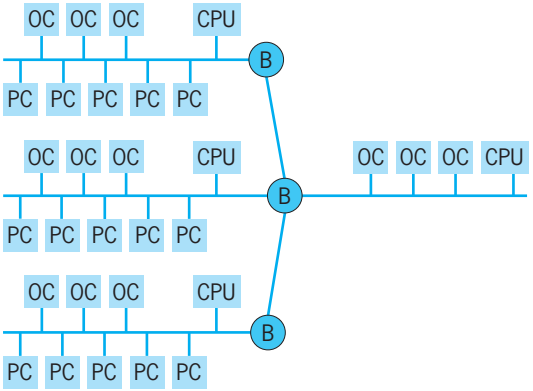

Fig. 4. Multihighway distributed system. PC = process or programmable controller module; OC = operator console; CPU = computer (central processing unit); B = bridge.

controllers permit the realization of more complex control algorithms than those available in comparable analog controllers, thereby retaining the advantage of the central direct digital control computer. The controller periodically transmits set-point, process signal, and controller status information to other modules for supervisory control purposes.

The success of direct digital control is dependent upon the periodic acquisition of the controlled variable, and this implies that the communication protocol must be able to provide this near-exact periodicity. Two protocols are used. First, the direct-digital-control application program may run periodically and, each time that data are needed, send a request message to the data acquisition devices on the network. These devices in turn acquire the data and return the value. In order for this to happen fast enough that the periodic control task in not disturbed, proprietary protocols are often used to tight-couple the devices and the control computer. For example, the request may hold the connecting bus until the device puts its data on the bus, thereby avoiding the overhead of a separate return message.

Second, a more recent protocol called Fieldbus uses a local microprocessor to handle a small number of data acquisition devices (100 or less) but allows the control computer to subscribe to the signal values at prespecified periodic intervals. The microprocessor then compensates for the communication delays by anticipating the need to send the data, and thereby initiates the data message without being requested each interval. This requires the control program to be carefully timed so that it is ready to receive and accept the data.

Operator consoles. An operator console usually has televisionlike screen displays and special functionally oriented keyboards which permit the process operator both to input commands to the control system which are equivalent to those input to integrated control systems, and to display values of process variables, groups of process variables, process status, and a variety of reports. In effect, the operator consoles function identically to those used with centralized digital control systems, except that any data required for a display or the result of any operator command must be transmitted to or from the appropriate modules in the distributed system.

Distributed control systems have expanded in size and scope. Operator consoles have undergone the greatest change. Operators must have available overviews of the entire plant and processes within the plant and be able to focus on sections of the processes to respond to alarms and other events. As a consequence, operator consoles are multiscreen-based devices, and the trend has been toward flexible organization of the screens and less dedication of individual screens to specific functions. A group of screens might be organized so that they behave like a single very large screen for overview purposes. This wall-like display facilitates cooperation among operators and managers. When the need arises to concentrate on an individual section of the process, one or more of the screens may be controlled by an operator to display this more specific information. *See* ELECTRONIC DISPLAY.

Operator consoles may display very large numbers of process variables, and hence the communication protocol usually cannot involve requesting values one at a time. Rather, the control usually subscribes to groups of data which are then passed to the console as a single large message. This group-subscription protocol is very effective in minimizing loading of the networks and thereby allowing timely acquisition of the data values.

Data logging, trending, and archiving. The display of a plot of a process variable over a previous time interval (minutes, hours, or days) is an important capability for a human operator. A data logging and trending module collects data for a number of points and stores it in a form which can be called up by an operator for display at any time. It is important to be able to archive such data for future uses, including planning operations, demonstrating correct operation of the process in defense of a legal claim, and operator training. The large volume of data to be collected makes it useful to assign this function to a separate module, often called a historian. Operator consoles must be able to recall archived data from the historian and display it in a flexible manner. Supervisory and other general-purpose computers also need access to the archived data.

Discrete programmable controllers. Discrete controls are appropriate in sequencing and batch process applications. The programmable controller is a device designed specifically for such applications and is most widely used in discrete manufacturing applications. The trend toward larger process control systems has led to the inclusion of discrete functions and the need for programmable controller functions. These may be implemented in a separate module or included as building-block algorithms in controller modules. Communication to this type of control module usually involves a block of data. Hence, communication is essentially the exchange of messages without critical timing requirements such as in periodic data acquisition. *See* PROGRAMMABLE CONTROLLERS.

Structure of nodes. In order to achieve the advantages of standard modules (that is, standard functions, standard programs, graceful degradation, and high reliability), it is convenient to modularize as much as possible and to restrict the functions carried out by each node to standard functions.

In a typical node (**Fig. 5**), a local communication system (usually a parallel bus) is provided which is independent of the serial communication links that tie nodes in the overall system together. Such a bus is very high speed (because it is physically short and because digital data are transferred in parallel) and separate so that another level of reliability can be built into the system. That is, individual modules tie to this bus, which in turn links to the external communication network by separate communication controllers. External links may be made redundant simply by providing more than one link and more than one communication controller per node.

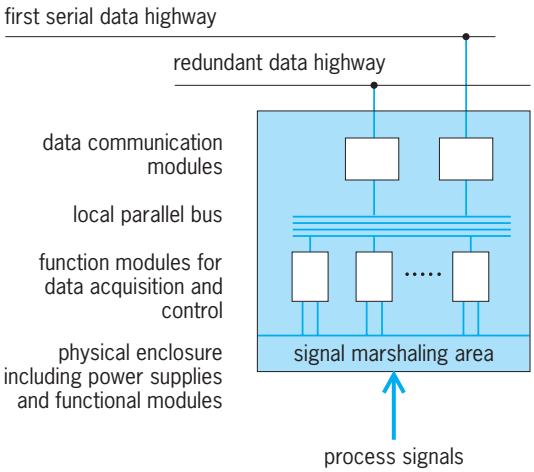

Fig. 5. Structure of one node in a distributed system showing internal functional modules.

Internally, functional modules are manufactured as separate printed circuit cards, and multiple cards are used to realize the required number of functions in a node. Hence failures on one printed circuit board do not necessarily destroy the function supplied by the node, nor does failure of one external communication link destroy the integrated system capability. Failure of all external communication links still leaves the functions within the node operating. Therefore graceful degradation can be extended to several levels. This capacity is extended even further by providing redundant power sources and power supplies.

Some manufacturers have even designed systems in which modules within a node are totally independent of one another from a functional point of view and interact only in that they use the internal local bus to communicate. An example might be a node which supports 32 control loops and is realized by four separate controller modules, each with its own input/output and power supply, and each communicating with the rest of the system via the local bus and the data communication modules, which themselves are independent and tie to the same local bus. Hence, many combinations of failures are possible which merely degrade performance rather than cause catastrophic breakdown.

Error detection and recovery within a controller module is often provided so that critical functions are not lost because of a single component failure. Error recovery requires switching to a backup module of the same type. It is necessary, however, that the current data being used by the control algorithms be available. Consequently, a node may provide a backup memory into which the process database in that node is periodically copied. After failure detection, the database is updated from the backup memory and the backup controller module is switched to replace the failed module. The process of snapshotting the process database and switching to backup hardware components, called fail-over, takes approximately 0.25 s from time of failure detection.

Serial data communication systems. The replacement of an integrated system by a set of interconnected modules does place one stringent requirement on a distributed system, namely that the communication network be fast enough to support all the demands of all modules for communication. For this reason, redundant communication links are often provided.

Data highways may be on the order of a mile (1.6 km) in length. Consequently, they are susceptible to errors produced by noise. Although the error rate is low, undetected errors cannot be tolerated because of the critical nature of the application. Hence all communication involves an exchange of messages specially formulated with check fields to ensure that all errors are detected. Errors are corrected by retransmission of the messages. The set of communication procedures and formats is termed a protocol.

Structure. The major distinguishing characteristic among communication protocols is the handling of multiple modules attempting to use the data highway simultaneously. The most commonly used protocols are ethernet (arbitrary initiation of message transmission when a quiet highway is detected, but stopping transmission when simultaneous transmission by two nodes is detected); active repeaters at each node (ring networks) so that simultaneous transmission may take place between each pair of nodes; and voluntary passing of mastership from node to node (token passing) in a logical ring structure. Although some smaller systems allow modules to communicate only in response to an invitation from a central computer (polling), this protocol does not provide adequate throughput and robustness for large systems. These differing philosophies lead to different response times to critical events.

Manufacturing application protocol. One communication protocol, the manufacturing application protocol (MAP), is being widely supported for control systems in discrete manufacturing. This standard attempts to make nodes and computers from different manufacturers compatible.

Multilevel systems. At the same time, there is increasing demand for very high levels of management in control applications. Examples include scheduling of generation loads, optimization of yield, and minimization of energy used for processing. These needs imply larger computers cooperating in the control process, and they are causing the structure of the modern control system to develop into a complex multilevel system (**Fig. 6**). Here the system consists of a set of processing units, each of which must be carefully monitored so that the overall plant control may be integrated. Distributed control systems are associated with individual process units. Each involves operator communication, supervisory control, and so forth.

A network of computers carrying out the overall plant high-level control couples the distributed systems together. The major distinction in the communication systems is that the higher-level computers do not acquire data or control physical process variables directly. Rather they communicate with the

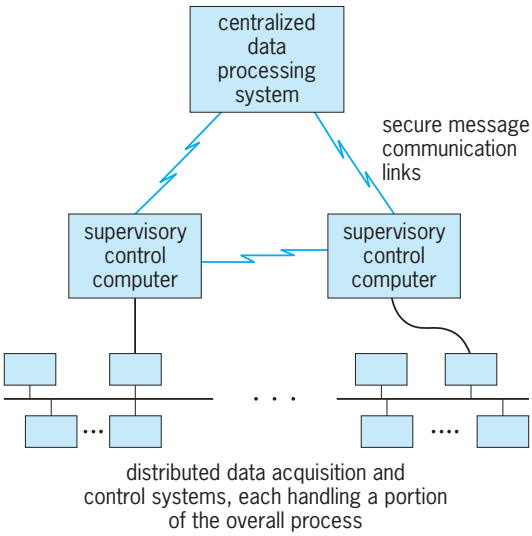

Fig. 6. Multilevel system with computers interconnected in a distributed network and using distributed data highway systems for direct communication with the process.

supervisory computers in the individual distributed control systems, exchanging messages to obtain data, parameter values, operating sequences, and so forth. *See* LOCAL-AREA NETWORKS.

Requirements for communication among the higher-level computers are very different from those for the data highways. Messages tend to be much longer (carrying a large amount of data) but require correspondingly slower response time. Hence such a network communication protocol is designed to be very general and error-free but not necessarily high-speed. This permits standards to be established which then result in low-cost hardware and software.

However, higher-level business-oriented applications need access to the real-time data collected and computed within the distributed control systems. Furthermore, large plants often have several different distributed control systems for different processes within the plants. The higher-level applications not only must access real-time data from different distributed control systems but also must combine the data in a higher-level application whose results are then used to manage the lower-level control systems. As a consequence, the preference is to require distributed control systems to be "open" rather than proprietary. Furthermore, the demand for real-time data necessitates a significant increase in data communication rates of distributed systems, requiring the use of new technology. Open standards for communication among the levels such as DCOM (distributed component object model) and CORBA (common object request broker architecture) are emerging.

Object-oriented concepts. Successful applications demand integration of data from a variety of sources and processing of the data in a flexible manner. These requirements impose severe demands upon software used in distributed control systems. In particular, this software must often be created specifically for the applications of a given plant or process. Object-oriented technology carries the promise of a building-block approach to the assembly of even very complex software. Objects are software artifacts specifically designed to encapsulate all the data and processing associated with a real-world entity. Control systems, process units, alarms, and other physically meaningful entities can be modeled as objects in software. Higher-level application software then uses these objects without being required to do the extensive software programming required to create them. This approach makes complex applications technically feasible to create and economical.

Standard interfaces. The success of object-oriented technology in distributed control systems is dependent upon standards for objects. These standard interfaces permit application software to use objects no matter where they reside in the system (physical distribution) or in what computer language they were programmed. They promise to make access to real-time data from distributed control systems as easily available as common applications on personal computers.

Object-based communication protocols. Object-based communication has emerged as a basis for communication between separate software modules either in the same computer or in separate computers. The DCOM and CORBA technologies allow the remote access of objects, called components. Each object contains some data, and some functions for manipulating that data. The overall system can be constructed by combining a given set of components, and may be updated and expanded by the addition of new components as the system needs them. The communication protocol views the module using a component no differently from an internal function. In fact, the data involved are packed into messages and transmitted across the underlying network in a standard manner which is opaque to the communicating modules.

A widely used example of object communication is the OPC (OLE for Process Control) specification, which defines a set of objects and their methods for performing data acquisition. In order to make data within a proprietary control system available to higher-level enterprise applications, a manufacturer may implement the OPC objects. The enterprise application then simply uses object-based communication to acquire needed data without any regard for communication protocols or the details of the implementation of the objects, which remains proprietary.

The combination of reliable, responsive distributed control and general-purpose communication networks leads to a system which can be adapted to critical control applications in a very flexible manner, with potential for increased productivity in plants, increased safety, and decreased energy consumption. Technology for higher-speed computation, data communication, and object-oriented software organization allows the integration of distributed control systems into plant-wide and enterprise-wise systems. *See* CONTROL SYSTEMS;

DIGITAL CONTROL; MULTILEVEL CONTROL THEORY; OBJECT-ORIENTED PROGRAMMING.

James D. Schoeffler

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Distribution (probability)

The results of a series of independent trials, random variables, or errors often occur in fairly regular and predictable patterns. These patterns can be expressed mathematically, and the most important are called the binomial, normal, and Poisson distributions.

Binomial distribution. Consider n independent trials, each of which results in success S or failure F, with corresponding probabilities p and $q = 1 - p$. Denote by S_n the number of successes. Because there are possible ways to select k places for S and $n - k$ places for F, the probability distribution of the random variable S_n is given by

$$\binom{n}{k}$$

where $k = 0, 1, \ldots, n$. This is the binomial distribution. Its expectation is np, its variance npq. *See* PROBABILITY.

If one lets a random variable X_k equal 1 or 0 according as the kth trial results in S or F, then $S_n = X_1 + \cdots + X_n$. The binomial distribution can therefore be approximated by the normal distribution in accordance with the central limit theorem. This special case is known as the DeMoivre-Laplace theorem; setting

$$x_k = (k - np)(npq)^{-1/2}$$

it asserts that

$$P\{S_n = k\} \approx (2\pi)^{-1/2} \exp(-x_k{}^2/2)$$

$$P\{a < S_n < b\} \approx (2\pi)^{-1/2} \int_{x_{a-1/2}}^{x_{b+1/2}} \exp(-t^2/2)\, dt$$

with a percentage error tending to 0 as $n \to \infty$ provided x_a and x_b are restricted to a fixed interval. (With $p = q = 1/2$ these formulas are useful for the evaluation of binomial coefficients and their sums.) *See* BINOMIAL THEOREM.

Normal distribution. The standard normal density is defined for $-\infty < x < \infty$ by

$$\phi(x) = (2\pi)^{-1/2} e^{-x^2/2}$$

The standard normal distribution function or error function $\Phi(x)$ is its integral from $-\infty$ to x. The normal distribution function with mean m and variance s^2 is

$$\Phi[(x - m)/s]$$

and its density is

$$(2\pi)^{1/2} s^{-1} e^{-(x-m)^2/(2s^2)}$$

As x goes from $-\infty$ to ∞, the function increases from 0 to 1. It plays an important role in many fields. In particular,

$$u(t, x; \xi) = 2\pi^{-1/2} t^{-1/2}\, e^{-(x-\xi)^2/4t}$$

is the fundamental solution of the heat (or diffusion) equation $u_t = u_{xx}$ and represents the heat distribution on the x axis at time t caused by a unit heat source initially concentrated at $x = \xi$. Probabilistically, this represents the transition probabilities in the Wiener process.

A random variable whose distribution is normal is called normal or gaussian. The sum of two independent gaussian variables is again gaussian; hereby the means and variances add. Analytically this means that the convolution of two normal distributions is again normal. This property characterizes the normal distributions among distributions with finite variances.

Central limit theorem and error theory. These are the best-known and most important applications of the normal distribution. The nature of the central limit theorem is best seen from the simplest special case: If $X_1, X_2, \ldots$ are independent random variables having a common distribution with mean m and variance s^2, their sum $S_n = X_1 + \cdots + X_n$ has mean $\mu = nm$ and variance $\sigma^2 = ns^2$; the corresponding standardized variable $S_n^* = (S_n - \mu)\sigma^{-1}$ has a probability distribution $P\{S_n^* \le x\}$ tending to $\Phi(x)$ as $n \to \infty$; in other words, the distribution of S_n gets close to $\Phi[(x - \mu)\sigma^{-1}]$.

The striking feature of this result is that an essential property of a sum of many independent components is independent of the nature of these components. The general central limit theorem shows this to be true under much wider conditions: The distribution of the sum S_n tends to normality even if the distributions of the components X_k vary with k, provided only that the components are likely to be of the same order of magnitude (so that no component has individually a noticeable effect on the sum). It is not even necessary that the X_k be independent.

Many empirical quantities (for example, the amount of water in a reservoir, certain inherited characteristics such as height, and the experimental error of physical measurements) represent the cumulative effect of many small components, and the statistical fluctuations of such quantities may be expected to follow the normal distribution. In particular, under the authority of K. F. Gauss, it has been assumed for a long time that experimental errors are approximately normally distributed. Modern research has shown the limitations of this assertion. Many distributions appear to the untrained eye as

nearly normal, but refined statistical tests prove that even the finest physical measurements depart considerably from normality, and that assignable causes (that is, large-contribution components) can be discovered statistically. This discovery shows the questionable character of the classical methods to predict the probable experimental error. In industrial quality control the departures from normality are used efficiently to discover assignable causes and thus to spot coming trouble at an early stage. *See* QUALITY CONTROL.

Misinterpretation of theory. Much harm has been caused by the widespread misinterpretation of the limit theorems in probability and of the meaning of statistical equilibrium in stochastic processes. The situation may be explained in terms of a coin-tossing game in which H and T count 1 and -1, respectively. Here $X_k = \pm 1$ with probability 1/2, and $m = 0, s^2 = 1$. The operational meaning of the central limit theorem in this case is as follows. Fix a large n. Repeat the same game of n tossings independently many times. One would then expect that in about 50% of the cases $S_n > 0$, in about 25% of the cases $S_n > 0.67n^{1/2}$, in about 16% of the cases $S_n > 2n^{1/2}$, and so on. What the central limit does not say is that in one game about half of the sums $S_1, \ldots, S_n$ will be positive. In fact, the arcsine law shows the opposite to be true: It is much more probable that all $S_j > 0$ than that there be equally many positive and negative ones.

Multivariate normal distribution. The previous theory carries over without essential changes to n dimensions. The n-dimensional normal density is defined by $(2\pi)^{-n/2}D^{1/2}e^{-Q(x_1,\ldots,x_n)/2}$, where Q is a positive definite quadratic form with determinant D; the matrix of variances and covariances is the reciprocal of the matrix of Q. If the n-dimensional joint distribution of the random variables $X_1, \ldots, X_n$ is normal, then each X_j is normal. The converse assertion is false, even though found in textbooks. The multivariate normal distribution is important for stationary stochastic processes. *See* STOCHASTIC PROCESS.

Poisson distribution. The Poisson distribution with parameter λ is the probability distribution of a random variable assuming the values 0, 1, 2, ... with probabilities $p_k(\lambda) = e^{-\lambda}\,\lambda^k/k!$. Both its expectation and variance equal λ. This is one of the most important distributions; it plays a basic role in the theory of stochastic processes and in many applications. A full understanding of its character can be obtained only from a postulational derivation and from the consideration of its many generalizations. However, much can be gained by the following elementary approach starting from the binomial distribution.

Consider a large number n of independent trials, each of which results in success or failure with probabilities p and $q = 1 - p$. Ordinarily, interest is restricted to the case where p is very small, but the expected number of successes $np = \lambda$ is of moderate size. Typical examples may be obtained by considering centenarians, color-blind people, or triplets in a large population; the defectives in an allotment of screws or fuses; or the wrong calls among all calls arriving during a day at a busy telephone. The number of successes in the n trials is a random variable with the binomial distribution, but under the present circumstances the binomial is close to the Poisson distribution and may be replaced by it. In fact, the probability of no success is $q^n = (1 - \lambda/n)^n$, which is close to $e^{-\lambda}$, the first term of the Poisson distribution. Now the ratio of the kth to the $(k - 1)$th term in the binomial distribution approaches that of the Poisson distribution.

This reasoning explains why the statistical fluctuations in the phenomena cited above follow approximately the Poisson distribution. In other circumstances the Poisson distribution appears, not as an approximation, but as the exact expression of a law of nature. This is true in particular of processes where certain events, such as radioactive disintegrations, mutations, power failures, and accidents, occur in time in such a way that (1) the probability that an event occurs during any given time interval of length dt is, asymptotically, λdt; and (2) there is no interaction or aftereffect between nonoverlapping time intervals. Under these assumptions a time interval of length t can be divided into $n = t/dt$ subintervals, each representing a trial in which the probability of success is close to λdt; a change to the limit $dt \to 0$ gives the Poisson expression $p_k(\lambda t)$ as exact probability for the occurrence of k events during a time interval of length t. A similar argument applies to random distributions of points in space, with t interpreted as volume; typical examples are stars, flaws of material, raisins in a cake, and animal litters in a field. Thus "perfect randomness" of a chance aggregate of points in space or time usually means that the fluctuations obey the Poisson law. *See* ANALYSIS OF VARIANCE; STATISTICS.

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District heating

The supply of heat, either in the form of steam or hot water, from a central source to a group of buildings. As an industry, district heating began in the early 1900s with distribution of exhaust steam, previously wasted to the atmosphere, from power plants located close to business and industrial districts. Advent of condensing-type electrical generating plants and development of long-distance electrical transmission led to concentration of electrical generation in large plants remote from business districts. Most district heating systems in the United States rely on separate steam generation facilities close to load centers. In some cities, notably New York, high-pressure district steam (over 120 lb/in.^{2} or 830 kilopascals) is used extensively to feed turbines that drive pumps and refrigerant compressors. Although some district heating plants serve detached residences, the cost

of underground piping and the small quantities of heating service required make this service generally unfeasible.

District heating, apart from utility-supplied systems in downtown areas of cities, is accepted as efficient practice in many prominent colleges, universities, and government complexes. Utility systems charge their customers by metering either the steam supply or the condensed steam returned from the heating surfaces.

District heating has grown faster in Europe than in the United States. A substantial part of the heat distribution is by hot water from plants combining power and heat generation. Heat from turbine exhaust or direct boiler steam is used to generate high-temperature (270–375°F or 132–191°C) water, often in cascade heat convertors, where live steam is mixed with return hot water. Hot water distribution systems have the advantage of high thermal mass (hence, storage for peak periods) and freedom from problems of condensate return. On the other hand, pumping costs are high and two mains are required, supply and return. (The cost of condensate return piping and pumping is so high that many steam distributors in the United States make no provision for condensate return at all.)

Reykjavik, Iceland, is completely heated from a central hot water system supplied from hot springs. Average water temperature is 170°F (77°C).

The most extensive European developments in district heating have been in Russia, mainly in the form of combination steam-electric plants serving new towns. The Russians developed a single-pipe high-temperature water district heating system. At each building served, superheated water from the main is diluted with return water from the building system. Excess water is used in kitchens, laundries, bathrooms, and swimming pools.

An important extension of the principle of district heating was the development of district heating and cooling plants. Usually owned and operated by the natural gas utilities, central plants distribute chilled water for air conditioning in parallel conduits with hot water or steam for heating. The system serving Hartford, Connecticut, is the pioneer installation. Because the costs of these four-pipe distribution systems are extremely high, development is slow and restricted to small areas of high-density usage within a few blocks of the plant. *See* COGENERATION; COMFORT HEATING; HOT-WATER HEATING SYSTEM.

Richard L. Koral

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Diving

Skin diving, or diving without the aid of special equipment, is a technique to investigate underwater environments and to gather shellfish or other items of commercial importance. In breath-hold diving, fins and faceplate are usually employed to facilitate the diver's activities. For longer periods of time, divers can function by using air pumped to them from the surface. This restricts the area that can be explored since movement of the diver is limited by the length of the air hose. Diving with scuba (self-contained underwater breathing apparatus) provides an almost limitless amount of freedom. With this equipment, the amount of time the diver can stay beneath the surface and the distance traveled are limited only by the volume of compressed air in the diver's tanks or aqua-lung.

Techniques

Humans have used compressed-air diving as a tool in marine research, salvage, and construction for over two and a half centuries. The history of diving is filled with crippling accidents and death, which are frequently the result of excessive pressure and abnormal gas ratios. One of the earliest records of diving and its problems is recorded on a plaque at the entrance to Old Port Royal, Jamaica, West Indies: "Lucas Barrett, F.G.S; L.L.S.—1837–1862—Died while conducting underwater research off Port Royal on December 19, 1862—25 years old—Geological Society of Jamaica."

Fortunately, efficient training programs, a much better understanding of diving physiology, and reliable equipment have removed many of the early dangers of diving. Diving is now an effective and safe means of conducting marine research. Although some scientific diving was accomplished prior to 1940, it was not until scuba (such as J. Y. Cousteau's aqualung) became commercially available in 1950 as inexpensive equipment for the sportsperson that diving became a widespread means of conducting marine research.

All modern methods of diving deliver a breathable gas mixture to the diver's lungs and body cavities at the same pressure as that exerted by the overlying water column. Thus there is no difference in pressure between the interior and exterior parts of the diver's body, and the diver has no feeling of pressure from the overlying water, regardless of depth. The method of supply and the duration of submergence varies with the type of equipment. Equipment can be divided into two basic types: free diving (scuba), where the diver carries a supply of compressed gas on the back; and tethered diving (hard hat), with a gas supply furnished to the diver via a hose from the surface.

Scuba diving. Scuba is extensively used by trained personnel as a tool for direct observation in marine research and underwater engineering. This equipment is designed to deliver through a demand-type regulator a breathable gas mixture at the same pressure as that exerted on the diver by the overlying water column. The gas which is breathed is carried in high-pressure cylinders (at starting pressures of 2000–3000 lb/in.^{2} or 14–21 megapascals) worn on the back. The scuba diver, free from surface tending, has much more maneuverability and freedom than the conventional hard-hat diver.

Fig. 1. A diver wearing a mixed-gas scuba, exposure suit, and face mask. Depth 300 ft (90 m) on the upper lip of Scripps Submarine Canyon, La Jolla, California. (*U.S. Navy photograph, Naval Electronics Laboratory Center*)

Fig. 2. Underwater scientific investigation by scuba diving, using compressed air. (*U.S. Navy photograph, Naval Electronics Laboratory Center*)

Scuba can be divided into three types: closed-circuit, semiclosed-circuit, and open-circuit. In the first two, which use pure oxygen or various combinations of oxygen, helium, and nitrogen, exhaled gas is retained and passed through a canister containing a carbon dioxide absorbent for purification. It is then recirculated to a bag worn by the diver (**Fig. 1**). During inhalation additional gas is supplied to the bag by various automatic devices from the high-pressure cylinders. These two types of equipment are much more efficient than the open-circuit system which does not take advantage of the unused oxygen in the exhaled gas. The exhaled gas is discharged directly into the water after breathing. The closed- and semiclosed-circuit types are, however, more complicated and can be dangerous if they malfunction or are used by inexperienced divers. Most open-circuit systems use compressed air because it is relatively inexpensive and easy to obtain. Although open-circuit scuba is not as efficient as the other types, it is preferred by most scientists and engineers who work underwater because of its safety, the ease in learning its use, and its relatively low cost.

Free diving or scuba diving (**Fig. 2**) using compressed air as the breathing gas is used to some extent in most current marine scientific and engineering investigations in all environments, including the Arctic. For physiological reasons the underwater observer is limited to about 150 ft (50 m) of water depth when obtaining observations which require thinking clearly and solving problems while submerged. Below this depth when using compressed air as a breathing gas, the diver is limited, not by equipment, but by the complex temporary changes which take place in the body chemistry while breathing gas (air) under high pressure. The practical depth limit to diving with air in clear, warm water is about 240 ft (80 m).

Tables containing permissible rates of ascent and time allowable at different depths are available from the U.S. Navy Experimental Diving Unit, Washington, D.C., and from the Manned Undersea Science and Technology Office of the National Oceanic and Atmospheric Administration (NOAA), Rockville, Maryland. Tables are also available through most commercial diving supply outlets. Breathing a mixture of helium, instead of nitrogen, with oxygen can increase the diver's ability to remain safely underwater for longer periods of time, but because of the expense and difficulties of handling the specialized diving gear and breathing mixtures needed for this type of deep diving, it is not practical for most scuba use.

By using scuba, scientists are no longer restricted to the sea surface but are able to investigate underwater problems at first hand. Free diving or scuba diving is used as a tool in offshore mapping of geologic formations, biological studies in ecology, underwater engineering, and detailed oceanographic measurements.

Saturation diving. Saturation diving permits long periods of submergence (1–2 weeks). It allows the diver to take advantage of the fact that at a given

Fig. 3. Sealab II, an underwater station for deep-sea divers; 57 ft (17 m) long, 12 ft (3.7 m) in diameter. (*U.S. Navy*)

depth the body will become fully saturated with the breathing gas and then, no matter how long the submergence period, the decompression time needed to return to the surface will not be increased. Using this method, the diver can live on the bottom and make detailed measurements and observations, and work with no ill effects. This type of diving requires longer periods of decompression in specially designed chambers to free the diver's body of the high concentration of breathing gas. Decompression times of days or weeks (the time increases with depth) are common on deep dives of over 180 ft (60 m). However, when compared with diving from the surface, this type of diving has been shown to be much more efficient since it permits approximately three times the amount of work to be done for the amount of time spent on a project. Furthermore, the marine fauna become accustomed to seeing the divers living among them, and behavioral studies are greatly improved. While saturation diving is an exciting method of diving that greatly extends the time spent underwater, the long periods of time needed to decompress and the limited number of underwater laboratories capable of supporting saturated divers have kept it from becoming a popular tool for the average scientist and engineer. Its primary use has been in the servicing of offshore oil-drilling rigs, in maintenance and repair of dams, and in scientific studies in research projects, such as Hydro-Lab, Tektite, and Sea Lab (**Fig. 3**). Factors such as cold water at depth, lack of light, the requirement for bulky and expensive equipment, and the presence of strong bottom currents place a practical limit to this type of diving at about 640 ft (195 m). *See* UNDERWATER VEHICLES.

Robert F. Dill

Physiology

Environmental effects on the submerged diver are quite different from those experienced at sea level. Two elements are very evident during the dive. As the depth of surrounding water increases, pressure on the air the diver breathes also increases. In addition, as the pressure increases, the solubility of the gases in the diver's tissues increases. The tissues, therefore, accumulate certain gases which are not metabolized. The increased presence of certain gases causes specific and often dangerous physiological effects.

Pressure. The total pressure of the atmosphere at sea level is approximately 760 mmHg or 30.4 in. Hg.

Pressure increases underwater at the rate of 1 atm (10 kilopascals) for each 33 ft (10 m) that the diver descends. The total pressure applied to the body and to the breathing gas increases proportionately with depth. As pressure of the gas increases, the amount of gas that is absorbed by the body increases. This is particularly evident if the diver is breathing air within a caisson since the percentage of nitrogen in the breathing gas increases proportionately to the amount of oxygen that is removed. Likewise, the amount of carbon dioxide in the body increases during the dive, particularly if the exhaled air is not separated from the inhaled air.

According to Dalton's law, the pressure exerted by an individual gas in a mixture (the partial pressure) is independent of the pressure of the other gases. The percentage or proportion of individual gases in air is always the same at sea level, during a dive, or at high altitudes. In other words, if the percentage of oxygen is 20.8% at sea level, it will also be 20.8% at any depth in the dive. However, if the diver is breathing air at a partial pressure of 100 mmHg (4 in. Hg) and descends 132 ft (40 m), which is approximately 4 atm of pressure, the partial pressure of oxygen increases to over 400 mmHg (16 in. Hg). The increase in pressure thus causes an increase in the effects of specific gases as well as exertion of a direct force on the body of the organism. In cases where the diver descends while wearing a helmet or scuba gear, the pressure within the diver's lungs must be maintained at a high level in order to prevent the pressure of the surrounding water from collapsing the thorax. In a caisson, the pressure exerted within the lungs is increased with increased depth. This is due to the increased pressure created by the surrounding water.

Effects of specific gases. One of the more obvious effects of gases on divers is caused by nitrogen. This gas makes up about 78% of the air that is normally breathed, and its solubility in the tissues increases as atmospheric pressure increases. When nitrogen is dissolved in the body, more than 50% is contained in the fatty tissues; this includes the myelin sheaths which surround many nerve cells. When divers undergo increased pressure, amounts of nitrogen in nerve tissue increase and lead to nitrogen narcosis or "rapture of the deep." Nitrogen narcosis can occur at 415 ft (130 m) or 5 atm (500 kPa), and increases in severity as the diver descends below this depth. The irrational behavior and euphoria often seen in nitrogen narcosis can result in serious, even deadly mistakes during a dive. The maximum time that divers can remain underwater without showing symptoms decreases with increasing depth. For instance, at 130 ft (40 m) the maximum time a diver can safely remain submerged in scuba gear is 15 min; in more shallow dives, the time becomes much longer. Obviously, nitrogen narcosis is one of the dangers of diving simply because the irrational behavior associated with it cannot be tolerated underwater.

One of the earlier recognized problems associated with human diving is known as decompression sickness, the bends, or caisson disease. If a diver is allowed to stay beneath the surface for long periods of time, the volume of dissolved gases in the tissues will increase. This is particularly true of nitrogen in the case of air breathing. When nitrogen accumulates in the tissues, it remains in solution as long as the pressure remains constant. However, when pressure decreases during the ascent, bubbles form in the tissues. Nitrogen bubbles can occur in nerves or muscles and cause pain, or they can occur in the spinal cord or brain and result in paralysis, dizziness, blindness, or even unconsciousness. Bubbles forming in the circulatory system result in air embolism. If the embolism occurs in the circulation of the lungs, a condition known as the chokes occurs. The first symptom noticed by the individual is pain beneath the sternum, followed by rapid shallow breathing; this ultimately results in pulmonary edema and can cause death. This condition does not occur if the nitrogen in the breathed air is replaced with hydrogen or helium. Because of its explosive qualities, hydrogen is avoided, but helium can be safely employed. Because of its increased mobility, helium moves out of the tissues rapidly and does not readily build up in them as bubbles. *See* DECOMPRESSION ILLNESS.

A method of prevention of decompression sickness was suggested by J. S. Haldane in 1907. Haldane introduced the method of stage decompression, in which the diver is allowed to ascend a few feet and then remain at this level until the gases in the tissues have been allowed to reequilibrate at the new pressure. This stepwise ascent is continued until the diver finally reaches the surface. A modern variation of this method consists of placing the diver in a decompression chamber after the surface is reached, to allow for periods of decompression which simulate ocean depths.

Oxygen toxicity is caused by increased concentrations of oxygen in the breathing gas. It can cause hallucinations, mental confusion, muscle twitching, nausea, and vertigo. When breathing pure oxygen, a diver is usually safe at 40 ft (12 m) for only 23 min and at 20 ft (6 m) for 1.5 h. The cause of oxygen toxicity is unknown. However, it leads to convulsions and coma, which are lethal if the diver is not brought to the surface immediately. The problem can be reversed once the diver reaches the surface and is given an adequate breathing mixture.

Carbon dioxide intoxication has not been a significant problem. However, in scuba diving it can occur occasionally. Carbon dioxide intoxication causes headaches, vertigo, and increased muscular activity. It has also been known to cause increased breathing rates and body warmth. Severe carbon dioxide intoxication can cause a diver to black out or lose consciousness even in shallow water; the depressant effects occur in the central nervous system without any previous warning signs. *See* RESPIRATION.

J. Homer Ferguson

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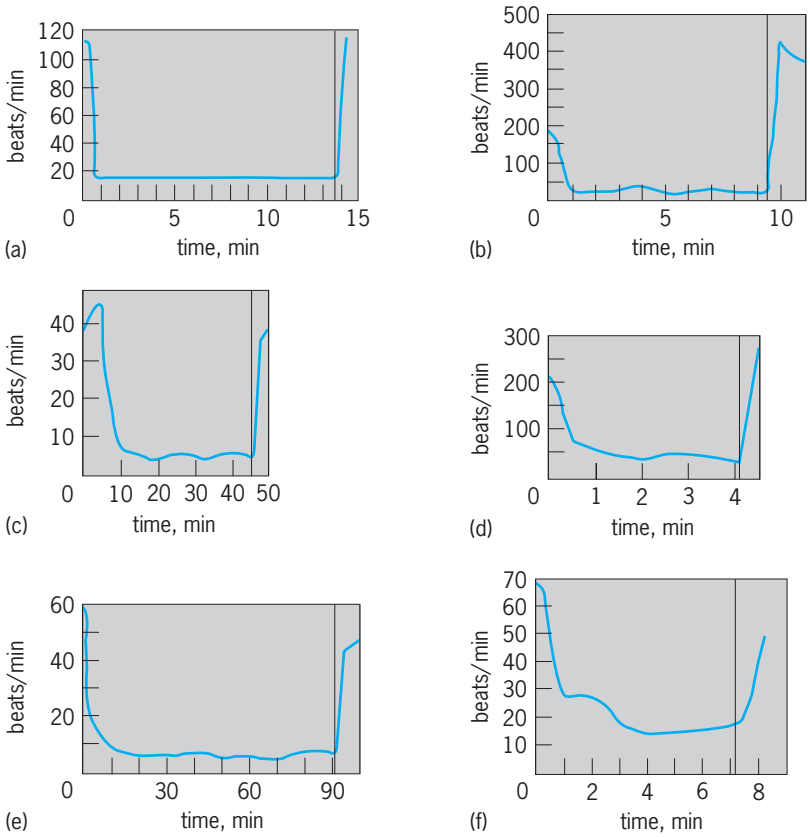
Diving animals

Various species of diving, air-breathing animals that can be found in every class of terrestrial vertebrates, having almost complete freedom in the aquatic habitat through several specialized physiological adaptations. Modern technology has allowed the investigation of many mechanisms that permit diving animals to spend long periods of time submerged. These investigations include studies to determine the physiological changes that occur during the dive. Truly aquatic animals have no need for a specialized diving response, because they are able to obtain oxygen directly from the water and thus sustain themselves indefinitely in the water. Terrestrial animals, on the other hand, have in many cases gone back in varying degrees to the aquatic mode of living. In all of these, the problem of supplying oxygen and nutrients to the tissues while submerged with only the physical attributes of a terrestrial animal has been a major specialization.

Physiological characteristics. Diving animals have been observed by scientists for many years. In 1870, Paul Bert measured the blood volume in birds and mammals relative to their breath-holding capabilities. He found that the blood volume in the bodies of these diving species was much greater than that found in nondiving animals. The larger proportion of blood to body volume gives these animals the ability to store larger amounts of oxygen. More oxygen allows the diving animals to remain submerged for longer periods of time than the nondiving species. The times and depths of dives of some species are shown in the **table**.

In many cases, animals undergo pronounced physiological changes when they dive. Above the surface of the water, they have all of the physiological characteristics of truly air-breathing organisms. When they dive, certain changes occur in their bodies which allow them to remain submerged for several minutes to an hour or more. One physiological change noted early was reduction in the rate at which the heart beats. This reduction, or bradycardia, is dramatic in the species which are well adapted to diving (see **illus.**). In the beaver, which has a normal resting heart rate of 140 beats per minute, the rate falls to less than 20 beats per minute after submergence.



Changes in heart rate in several species during diving: (*a*) hair seal; (*b*) duck; (*c*) alligator; (*d*) muskrat; (*e*) common water snake; and (*f*) common garter snake.

Duration and depth of diving reported for various species

Species	Duration, min	Depth, ft (m)
Mink *(Mustela vision)*	3	—
Muskrat *(Ondatra zibethicus)*	12	—
Walrus *(Odobenus rosmarus)*	10	—
Gray seal *(Halichoerus grypus)*	20	330 (100)
Sperm whale *(Physeter catodon)*	75	2950 (900)
Common garter snake *(Thamnophis sirtalis)*: forced dive	7	—
Common water snake *(Natrix sipedon)*: forced dive	90	—

Bradycardia is the result of a specific autonomic reflex which controls heart rate as well as the flow of blood to the muscles and internal organs. During the dive, blood vessels that supply skeletal muscles, kidneys, viscera, and the periphery of the body are constricted. Blood flows occurs primarily to the brain, lungs, and heart muscle while the diving animal is submerged, and blood flow to the peripheral organs is reduced or completely restricted. In this condition the animals derive energy solely from stored metabolites and oxygen. This reflex prevents lactic acid (lactate) from entering the circulation. Lactic acid is a metabolic by-product of the Embden-Meyerhof pathway, or anaerobic glycolysis. Since lactate remains in the tissues, the likelihood of its stimulating chemoreceptors found in the larger blood vessels is reduced. *See* CARBOHYDRATE METABOLISM; CHEMORECEPTION.

Another characteristic of diving animals is the capacity to store large quantities of glycogen, adenosine triphosphate (ATP), and oxygen in their muscles. Much of the oxygen used by the peripheral organs of the body is stored by means of a compound

called myoglobin. Myoglobin is higher in concentration in the muscles of diving animals than in those of the nondiving ones. *See* MUSCLE PROTEINS; MUSCULAR SYSTEM.

Some diving animals have an enhanced ability to generate energy via anaerobic glycolysis. The enzymes involved in anaerobic glycolysis, such as lactic acid dehydrogenase and phosphokinase, are in greater concentration in the organs of diving animals than in the organs of nondiving species. Since they have greater amounts of these enzymes in their tissues, diving animals can carry out anaerobic glycolysis much more rapidly and efficiently than the nondiving species.

Oxygen debt develops when an animal utilizes glucose anaerobically. When this occurs, glucose is broken down to pyruvic acid through several intermediate stages which yield energy in the form of ATP. Pyruvic acid (pyruvate) is converted to lactic acid in the absence of oxygen. Lactate can then be converted back to pyruvate when oxygen is available. The accumulation of lactic acid in the tissues is associated with oxygen debt. When oxygen is obtainable again through breathing, the animal is said to repay its oxygen debt. Lactic acid dehydrogenase is an enzyme which is essential for the conversion of pyruvic acid to lactic acid in the absence of oxygen. It is also capable of catalyzing the reverse of this reaction when oxygen is supplied to the tissues. The presence of elevated concentrations of lactic acid dehydrogenase in the organs of diving animals allows them to repay oxygen debt more rapidly after surfacing.

All of the chemical reactions carried out in an organism constitute its metabolism. Where metabolism can be reduced, the dependence upon oxygen can also be reduced. It has been shown that in diving animals the metabolic rate is slowed when they are submerged. This is indicated by a decrease in body temperature which is 4–6°F (2–3°C) below normal during the dive. Such a drop in body temperature would result from a 25% decrese in total metabolism.

During the dive, insulation of the body core is increased as the result of a reduction in blood flow to the periphery. This effectively reduces the amount of heat exchanged between the core of the body and the environment. This is essential in reducing total metabolism. *See* CARDIOVASCULAR SYSTEM.

Diving species. The diving response has been shown in amphibians, reptiles, birds, and mammals and consists of several complex physiological changes which allow the animals to remain submerged for long periods of time.

Mammals. Among the mammals, the mink and otters, members of the family Mustelidae, are excellent divers but normally do not remain beneath the surface for more than a few minutes. The muskrat (*Ondatra zibethicus*) has been known to remain submerged for up to 12 min; however, the depth to which it dives is not definitely known. Another excellent and well-known fresh-water diver among the mammals is the beaver (*Castor canadensis*). The better divers among mammals are found in the seals, which are usually located in the colder waters of the Arctic and Antarctic. The Weddell seal, which is found in the Antarctic, is a champion among divers, going to depths of 2000 ft (600 m) and remaining submerged for up to 43 min. This is quite remarkable since the pressure of the water increases up to 883 lb/in.^{2} (609 kilopascals) at this depth. In order to sustain a dive at this depth, the organism must be able to withstand temperatures as low as 37°F (2°C). The seals are especially useful in diving experiments. In addition to their excellent diving capabilities, they are readily adaptable to captivity and can be trained to dive at the investigator's direction.

Of all the aquatic and marine mammals, whales are probably the best adapted to diving. The sperm whale (*Physeter catadon*) has been known to remain submerged for as long as 75 min at a depth of 2950 ft (900 m). The blue whale *(Sibbaldus musculus)* can dive for 49 min and reaches a depth of 330 ft (100 m). One of the factors which contributes to the depths to which whales can dive is related to their ability to collapse their lungs. Whales force the air expelled from the lungs into the trachea, which is expanded in these species and which is impermeable enough to prevent gaseous exchange with the tissues. This eliminates the possibility of absorbing high concentrations of nitrogen from the air in the lungs during the dive. Whales also have phenomenal thermoregulatory mechanisms which allow them to exist in both arctic and temperate-zone waters. *See* MAMMALIA.

Birds. A wide variety of birds are known to be excellent divers. This includes birds such as the dipper and the osprey, which are submerged for only a short period of time while feeding. On the other hand, many bird species such as the penguin can remain submerged for a long period of time. They have lost the ability to fly and are well adapted to the aquatic environment. Almost all of the ducks and geese are divers and can remain submerged from a few minutes to nearly 20 min. This group includes the grebes and loons, which are web-footed and have short wings that enable them to be excellent swimmers. *See* AVES.

Reptiles. Many reptiles are also excellent divers. During the voyage of the *Beagle*, Darwin observed the diving iguanas which live on the Galápagos Islands, off the coast of Ecuador. These lizards spend much of their lives in the aquatic habitat. Diving reptiles are also found among snake species. These include the completely marine snake (*Hydrophis belcheri*) and those which are only partially aquatic, such as the common garter snake *(Thamnophis sirtalis)* and the common water snake *(Natrix sipedon)*. Other divers among the reptiles are the alligator, crocodile, and cayman. These animals can remain submerged for prolonged periods of time partially because of their lower metabolic rate and thus lower demand for respiratory oxygen. Other good examples of divers in the class Reptilia include the turtles, both aquatic and marine. These animals may stay submerged for hours or even days depending upon the season. *See* REPTILIA.

Amphibians. The amphibians are also excellent divers and can remain submerged for long periods. All amphibians should be included in the category of divers, since they are all partially aquatic and must become submerged from time to time in their life cycle. *See* AMPHIBIA; DIVING. J. Homer Ferguson

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Division

In arithmetic and algebra, the process of finding one of two factors of a number (or polynomial) when their product and one of the factors are given. The symbol ÷ now used mostly in elementary English and American arithmetics to denote division first appeared in print in an algebra by J. H. Rahn published in Zurich in 1659. Division is more often symbolized by the double dot :, the bar —, or the solidus /; thus $x:y$, or x/y indicates division of a number x by a number y. Considered as an operation inverse to multiplication, x/y is a symbol denoting a number whose product with y is x. Another way to base division upon multiplication is provided by the concept of the reciprocal of a number. If y is any number (real or complex) other than 0, there is a number, denoted by $1/y$ and called the reciprocal of y, whose product with y is 1. Then x/y is the symbol for the product of x and $1/y$. This view of division furnishes a means of extending the concept to objects other than real or complex numbers. A whole number is divisible by 2 if its last digit is so divisible, and by 4 if the number formed by the last two digits is so divisible. It is divisible by 3 or 9 according to whether the sum of its digits is thus divisible, respectively, and is divisible by 11 if the difference between the sum of the digits in the odd and the even places can be so divided. *See* ADDITION; ALGEBRA; MULTIPLICATION; NUMBERING SYSTEMS; NUMBER THEORY; SUBTRACTION. Leonard M. Blumenthal

DNA helicase

Any of a ubiquitous class of enzymes that catalyze the unwinding of double-stranded deoxyribonucleic acid (dsDNA) into single-stranded DNA (ssDNA). In all organisms (other than some viruses), genetic information is locked within a double helix formed by two antiparallel DNA strands. Although dsDNA is suitable for secure information storage, hydrogen bonds formed between complementary bases (Watson-Crick base pairing) impair readout of this information by the cellular machineries that frequently require ssDNA as the template. The unwinding of dsDNA into ssDNA that is catalyzed by DNA helicases is a function critical for virtually every aspect of cellular DNA metabolism including replication, recombination, and repair.

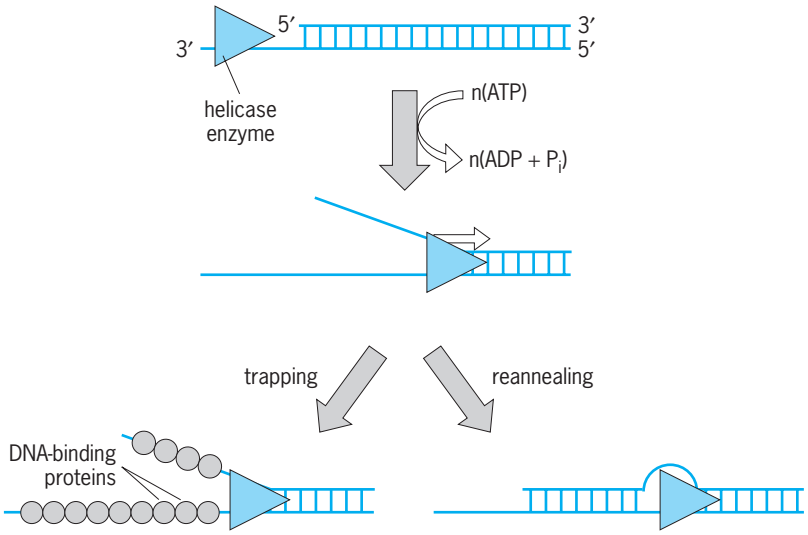

Fig. 1. Schematic representation of DNA helicase action. The helicase enzyme translocates along the DNA molecule and separates the strands. Energy for this unfavorable reaction is coupled to the hydrolysis of adenosine triphosphate (ATP) to adenosine diphosphate (ADP) and inorganic phosphate (P_i). In the presence of a single-stranded DNA-binding protein the nascent ssDNA strands are trapped and reannealing of the DNA duplex is prevented. The helicase depicted here displays a 3′ → 5′ polarity, tracking unidirectionally along the lower of the two DNA strands in the duplex (the loading strand).

First identified in the 1970s, DNA helicases are motor proteins that convert chemical energy into mechanical work. Chemical energy is derived from the hydrolysis of adenosine triphosphate (ATP) or other nucleoside triphosphates, and is coupled to mechanical work during at least two important steps within the helicase reaction cycle (**Fig. 1**): (1) the unidirectional translocation (movement) of the enzyme along the DNA molecule, and (2) the separation (unwinding) of the DNA duplex (that is, the disruption of hydrogen bonds between base pairs). *See* ADENOSINE TRIPHOSPHATE (ATP); DEOXYRIBONUCLEIC ACID (DNA).

Classifications. On the basis of amino acid sequence comparisons, helicases are divided into three large superfamilies and two smaller families that are illustrated schematically in **Fig. 2**. Moreover, these analyses identified several protein signature motifs (short conserved regions in a protein sequence) that are indicative of nucleic acid strand-separation activity and/or the capacity to translocate along DNA. Biochemical and structural data show that helicases can function as monomers, dimers, and multimers (predominantly hexamers) and that they can be classified on the basis of their substrate requirement for dsDNA, double-stranded ribonucleic acid (dsRNA), or DNA-RNA hybrids. Helicases are further classified as either 3′ → 5′ or 5′ → 3′ enzymes based on their polarity of unwinding (a region of ssDNA of the given polarity must be adjacent to the dsDNA to be unwound, to act as a loading site for the helicase).

Superfamilies 1 and 2. Superfamilies 1 (SF1) and 2 (SF2) include thousands of DNA and RNA helicases. SF1 and SF2 enzymes are primarily monomeric and dimeric helicases that unwind DNA with either 3 → 5′ or 5′ → 3′ polarity. Domains 1A and 2A in SF1 helicases and domains 1 and 2 in SF2 helicases contain

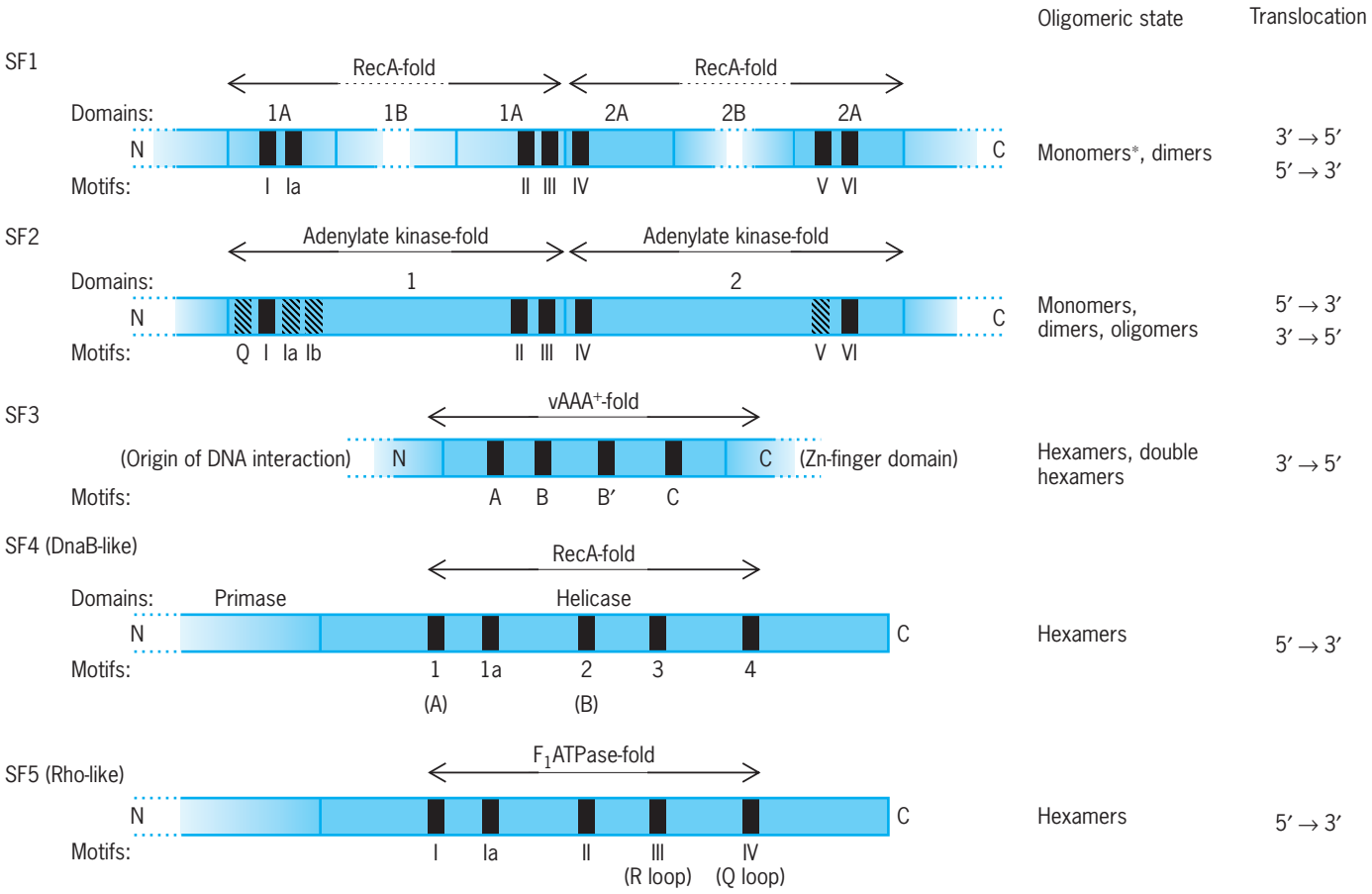

Fig. 2. Five major superfamilies of DNA helicases. The conserved signature amino acid sequence motifs are shown relative to the modular domains (shown as solid blocks) of helicases that belong to five major families. The motifs that are characteristic of all helicases within a particular superfamily are shown in black, while the motifs characteristic only to subsets of the SF2 helicases are represented by a hashed block. Alternative designations for some signature motifs are given in parentheses. In addition to the core domains containing the signature motifs, helicases frequently contain domains either adjacent to or inserted between the conserved motifs of the core domain. Examples of such domains are shown as discontinuous light color blocks. The areas containing large insertions that define additional enzymatic activities are indicated by broken lines. The oligomeric state and translocation polarity for each superfamily is indicated to the right of the block diagram.

similar sets of seven helicase signature motifs that comprise the basic structural core of these enzymes (Fig. 2). This core is responsible for both nucleotide binding and hydrolysis, and it includes a repeat of either of two structural domains: the RecA-fold in SF1 helicases or the related adenylate kinase-fold in SF2 helicases. Variation at the N and C termini, as well as the insertions between the core domains, give rise to differences in the substrate specificity or to additional enzyme activities. Motifs I and II [Walker A (GxGKT) and B (DExo/H) motifs, respectively] are the most conserved among all helicases.

Superfamily 3. Superfamily 3 (SF3) helicases are found in small DNA and RNA viruses. In contrast to SF1 and SF2 helicases, each subunit of these enzymes is composed of a single *vAAA*$^+$-fold that is responsible for nucleotide binding and hydrolysis.

Superfamilies 4 and 5. The smaller superfamilies 4 and 5 consist primarily of 5′ → 3′ hexameric helicases related either to the *Escherichia coli* DnaB helicase (SF4) or to the transcription termination factor Rho (SF5). In addition to Walker A and B motifs (characteristic of the RecA-fold), SF4 helicases contain three distinct conserved motifs (designated as motifs 1a, 3, and 4). All known SF4 enzymes are involved in the replication of bacteria and bacteriophages. These helicases are functionally and physically associated with DNA primases (enzymes that initiate the synthesis of each Okazaki fragment, short segments of newly synthesized DNA, by first synthesizing a short segment of RNA) either as two distinct but interacting proteins (for example, bacterial DnaB helicase and DnaG primase), or as two domains of one polypeptide, such as gp4 helicase-primase of bacteriophages T7 and T3 (as shown in Fig. 2). The nucleotide-binding and hydrolysis core of F5 helicases possesses the highest degree of similarity to the F1-ATPase fold, which is characteristic of proton-translocating ATPases. *See* ENZYME.

Directional translocation and DNA unwinding. It is now widely accepted that the unwinding polarity of helicases relates to a directional bias in translocation on ssDNA. For example, the enzyme depicted in Fig. 1 is a 3′ → 5′ helicase. Upon binding to the ssDNA, it starts moving toward the 5′ end of the loading strand, which brings the enzyme to the ssDNA-dsDNA junction; subsequent translocation along the ssDNA, coupled to either physical or electrostatic

destabilization of the dsDNA, results in strand separation in the duplex portion of the substrate. Conversely, a $5' \rightarrow 3'$ helicase translocates toward the $3'$ end of the loading strand and, therefore, requires a $5'$-terminated region of ssDNA to unwind a DNA duplex.

Interestingly, high-resolution structural data suggest that the helicase signature motifs are not essential for the duplex DNA separation per se, but are for the ATP-dependent unidirectional motion of the helicases on the DNA lattice. Consequently, it is proposed that the helicase signature motifs define a modular structure for a DNA motor, while the additional domains, which may vary from one protein to another, are responsible for modulating DNA unwinding, mediating protein-protein interactions, or determining structure specificity. Although the classical substrate for a DNA helicase is duplex DNA adjacent to a region of ssDNA (as depicted in Fig. 1), some enzymes preferentially unwind more complicated structures, such as flaps, forks, bubbles, or three- and four-way junctions.

DNA translocation can also be characterized in terms of processivity, which is defined as the probability of forward translocation at each base pair opening cycle. For a highly processive enzyme, stepping forward is much more likely than dissociation. Consequently, an enzyme that never dissociates after each step has a processivity equal to one (which is a physically unrealistic situation in which the enzyme is infinitely processive), whereas an enzyme that dissociates after each step has a processivity equal to zero (and is referred to as being dispersive or distributive). The average number of base pairs unwound (or translocated) by an enzyme in a single binding event (N) relates to the processivity (P) as $N = 1/(1 - P)$. For example, the UvrD helicase, whose processivity is 0.98, unwinds on average 50 base pairs. On the other hand, the highly processive RecBCD, a multifunctional enzyme involved in the repair of dsDNA breaks by homologous recombinations, has a processivity of 0.99997 and unwinds on average 30,000 base pairs before dissociation.

Step size. Though seemingly simple, one of the more poorly defined parameters for DNA helicase behavior is the "step size." The step size has been defined in at least three different ways: (1) as the number of base pairs that the enzyme unwinds upon hydrolyzing one ATP molecule under its most efficient reaction conditions; (2) as a physical step size, which is a measure of the helicase's ability to "step over" a discontinuity in the DNA; and (3) as a distance associated with a kinetically rate-limiting step. The parameters obtained using these three definitions are not necessarily identical. For example, the RecBCD enzyme hydrolyzes about two molecules of ATP per base pair unwound, but has a kinetic step size of six base pairs. The physical step size of the related RecBC enzyme was determined to be 23 base pairs.

Accessory factors. Once a helicase has separated the strands of dsDNA, spontaneous reannealing of the duplex can be avoided if the nascent ssDNA strands are trapped by an ssDNA-binding protein or by other factors that pass the unwound nucleic acid intermediate to the next step of a reaction pathway (Fig. 1). While ssDNA-binding proteins have frequently been shown to stimulate helicase activity in vitro, helicase activity is also stimulated by other accessory factors that increase the rate or processivity of unwinding (such as processivity factors, clamp loaders, and polymerases). The primary replicative helicase of *E. coli*, DnaB, is a good example of a helicase that acts poorly in isolation but, as part of the replisome (the DNA synthesis machinery of the cell), efficiently separates the DNA strands at the replication fork. The rate of movement of the replication machinery at the fork is coordinated by an interaction between DnaB and DNA polymerase (the enzyme that synthesizes the complementary strand of DNA) that enhances its translocation rate by almost 30-fold to 1000 base pairs per second.

The physical arrangement and interactions between proteins comprising a replisome modulate their respective activities to ensure complete and coordinated synthesis of both DNA strands. DnaB is a hexameric helicase that translocates in a $5' \rightarrow 3'$ direction on the DNA strand used as a template for lagging-strand synthesis. For its function within a replisome, DnaB helicase must be loaded on the lagging-strand template. This task is accomplished by the primasome assembly proteins, which coordinate recruitment of DnaB during establishment of the replication fork. Besides DNA polymerase and primasome assembly proteins, DnaB interacts with DnaG primase.

Single-molecule translocation visualization. Until recently, virtually all biochemical data on helicases were derived from conventional bulk-phase techniques, which observe the population-averaged properties of molecular ensembles. Direct observation of dsDNA unwinding by a single helicase enzyme has the potential to uncover valuable information on protein and DNA dynamics that may be obscured in bulk-phase assays. Two of these techniques were used to visualize translocation of a single molecule of the exceptionally fast and processive helicase RecBCD, which can unwind, on average, 30,000 bp of dsDNA per binding event, at a rate of 1000 bp/s, while simultaneously degrading the ssDNA products of its helicase activity.

In one approach, a device called an optical trap is used to manipulate individual, fluorescently labeled DNA molecules and to visualize their unwinding and degradation by RecBCD enzyme. In the presence of ATP, RecBCD enzyme unwinds the dsDNA, which is observed as a progressive shortening of the fluorescently labeled DNA molecule. An alternative single-molecule approach is the tethered particle motion experiment, which uses light microscopy to follow translocation of a biotin-tagged RecBCD enzyme bound to a streptavidin-coated polystyrene bead. The two single-molecule experiments are different yet complementary: the tethered particle motion experiment directly measures translocation,

whereas the optical trap method measures dsDNA unwinding. Together, these studies provide additional strong evidence for the coupling of DNA strand separation with movement of the helicase protein on its substrate lattice. Both single-molecule visualization methods show that RecBCD translocates unidirectionally and processively on dsDNA, with each molecule moving at a constant rate (within the limit of experimental detection). Although the average translocation rate is similar to that derived from bulk measurements, considerable variation is observed in the translocation rate of individual RecBCD enzymes.

These and other new techniques of single molecule detection will play an important role in determining the molecular mechanisms of DNA helicases. The next big challenges in this field are to understand how these DNA motors are incorporated into and function within the context of macromolecular machines to orchestrate complex DNA processing events. *See* MOLECULAR BIOLOGY; NUCLEOPROTEIN.

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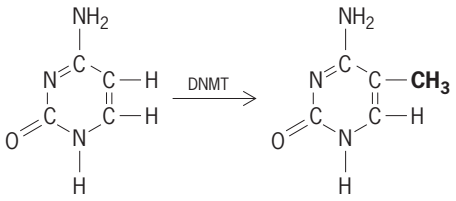
DNA methylation

The addition of a methyl group (—CH_3) to certain nucleotides in genomic deoxyribonucleic acid (DNA), a process which has important effects on gene expression. Besides adenine, guanine, cytosine, and thymidine, the four bases which spell out the primary sequence of genomic DNA, methylation produces a fifth base called methylcytosine. Methylcytosine is a covalent modification of DNA that is replicated when the cell divides. The enzymes that can add a methyl group to the cytosine ring to form methylcytosine are called DNA methyltransferases (DNMTs) [see reaction below]. In humans and other mammals, this



Cytosine 5-Methylcytosine

modification is imposed only on cytosines that precede a guanosine in the DNA sequence (the CpG nucleotide). In simple organisms, such as yeast and *Drosophila*, there is little or no DNA methylation. *See* DEOXYRIBONUCLEIC ACID (DNA).

The bulk of methylation in human DNA occurs in the large portion of the DNA that does not encode genes. This heavily methylated DNA replicates later than nonmethylated DNA. Late replication is associated with the formation of inactive chromatin (the threadlike strands of DNA complexed with histone proteins that form eukaryotic chromosomes), which facilitates transcriptional silencing of noncoding regions such as inserted viral sequences, repetitive DNA elements, and transposons (DNA sequences that move from their usual location into a new region of the genome). *See* NUCLEIC ACID.

Effect on gene expression. Studies in mammals have linked patterns of DNA methylation to gene expression. CpG dinucleotides are often clustered in small stretches of DNA termed CpG islands. Almost half the genes in the human genome have CpG-rich promoter regions, stretches of DNA where transcription of DNA into ribonucleic acid (RNA) begins. Methylation in a gene promoter region generally correlates with a silenced gene.

DNA methylation by itself does not directly repress transcription. Only after proteins are added to methylated sites in DNA, thereby forming chromatin, does transcriptional repression occur. Nucleosomes are structures composed of a core of histone proteins around which DNA winds. Long sequences of nucleosomes coil to form chromatin. Normally, at sites of active genes with nonmethylated CpG islands in promoter regions, the nucleosomes are widely and irregularly spaced in a manner favorable to the access of proteins that activate transcription. In contrast, when CpG islands are heavily methylated, the nucleosomes are tightly compacted and regularly spaced. This configuration excludes proteins that activate gene transcription or prevents them from acting as gene activators. Another regulatory feature of nucleosomes is contained in certain amino acid residues of histone tails. The covalent attachment of acetyl groups to lysine residues (acetylated state) in histone tails decreases the interactions between nucleosomes within the chromatin fiber, resulting in greater access of transcription factors to the DNA. These tails are kept in an acetylated state in the case of transcribed genes but in a deacetylated state in the case of hypermethylated, silenced genes. *See* GENE; GENE ACTION; NUCLEOSOME.

Role in cancer. Changes in the pattern of DNA methylation are commonly seen in human tumors. Both genomewide hypomethylation (insufficient methylation) and region-specific hypermethylation (excessive methylation) have been suggested to play a role in carcinogenesis.

DNA hypomethylation. The global level of DNA methylation is generally lower in tumor cells than in normal cells. This decrease in the overall level of DNA methylation is curious in light of the normal to high levels of DNA methyltransferase expression usually seen in tumor cells. The biological significance of DNA hypomethylation in cancer is not well understood. Because both genomic instability and hypomethylation are observed early in tumorigenesis, it has been tempting to speculate that genomic hypomethylation provides the incipient cancer cells with a way of acquiring more mutations by destabilizing the

genome. Defects in DNA methylation might contribute to chromosomal instability observed in aneuploid human colorectal cancers (that is, colorectal cancer resulting from the addition or deletion of whole chromosomes). Also, DNA hypomethylation of very discrete locations within the genome has been associated with abnormal chromosomal structures as observed in cells from patients with ICF (immunodeficiency, centromeric instability, and facial abnormalities) syndrome. This rare recessive disease is caused by mutations in the catalytic domain of the gene encoding one of the three biologically active DNMTs. *See* CHROMOSOME; HUMAN GENETICS; MUTATION; TUMOR.

DNA hypermethylation. DNA hypermethylation of CpG islands in gene promoter regions is associated with the inappropriate transcriptional silencing of tumor-suppressor genes, which indicates the important and pervasive role of DNA hypermethylation in oncogenesis. Because of its gene-silencing effect, abnormal promoter hypermethylation can have the same effect as a coding-region mutation in a copy of the gene. Examples are the *VHL* gene in renal cancer, the cell-cycle-control gene *p16* in several cancer types, and the mismatch-repair gene *hMLH1* in a subtype of colorectal cancers.

DNMTs. There are three known biologically active DNMTs in mammalian cells that catalyze the methylation process: DNMT1, DNMT3a, and DNMT3b. It is certain that each of these proteins is vital for embryonic development because inactivation of the corresponding genes in mice causes embryonic or early postnatal death. But much remains to be determined about how each participates in establishing normal DNA methylation patterns and the aberrant patterns in cancer. Though DNMT1 accounts for most methylation in mouse cells, human colorectal cancer cells lacking the *Dnmt1* gene retain significant genomic methylation and associated gene silencing. And disruption of the human *Dnmt3b* gene has been shown to reduce global DNA methylation by less than 3%. Surprisingly, genetic inactivation of both *Dnmt1* and *Dnmt3b* was found to nearly eliminate methyltransferase activity and reduce genomic DNA methylation by greater than 95%. These marked changes resulted in demethylation of repeated sequences, loss of insulinlike growth factor II (IGF2) imprinting (the differential expression of a gene depending on whether it was maternally or paternally inherited), abrogation or silencing of the tumor suppressor gene *p16INK4a*, and growth suppression. (IGF2 is thought to be an important growth factor in tumorigenesis; however, the specific physiologic effects of IGF2 in cancer are not known.) These results provide compelling evidence that the two enzymes cooperatively maintain DNA methylation and gene silencing in human colorectal cancer, and that such methylation is essential for optimal neoplastic proliferation. *See* ENZYME.

Diagnostic and therapeutic implications. The potentially reversible epigenetic changes in neoplasia present new opportunities for the clinical management of cancer. Strategies could potentially be developed that reverse gene silencing for the purpose of preventing and treating tumors. Moreover, the molecular changes associated with gene silencing may serve as markers for risk assessment, diagnosis, and prognosis. *See* CANCER (MEDICINE).

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Docodonta

A group of early mammals that currently includes at least eight genera from Laurasia (an ancient northern supercontinent comprising the landmasses that are now North America, Europe, Asia, Greenland, and Iceland) ranging in age from Middle Jurassic (159–180 mya) to Early Cretaceous (94–144 mya). Genera from the Late Triassic of Europe, Early Jurassic of India, and Late Cretaceous of South America have been tentatively referred to the Docodonta, but their evolutionary relationships are uncertain. All known docodonts were small. The skull of the most completely known docodont, the Late Jurassic genus *Haldanodon,* is approximately 30 mm (1.2 in) in length (**Fig. 1**). A reconstruction of its skeleton indicates the animal was equivalent in size to a modern mole. Other docodonts, known from isolated teeth or, at best, fragments of jaws and skulls, were of a similar size. *See* CRETACEOUS; JURASSIC.

The molariform cheek teeth of docodonts are transversely widened with two anteroposteriorly oriented rows of cusps linked by transverse crests (**Fig. 2**). Occlusion of the molariforms produced both a puncturing, or cutting, action and facilitated crushing and possibly grinding of their food, which probably included insects and other invertebrates. In

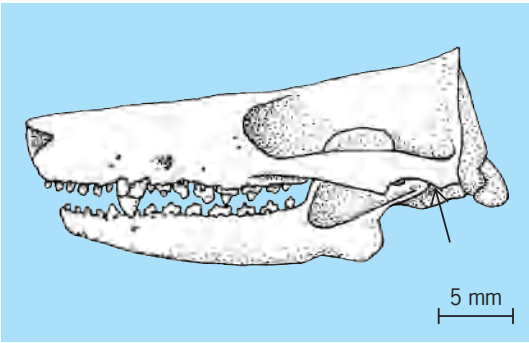

Fig. 1. Restoration of the skull of the Late Jurassic docodont *Haldanodon* in lateral view. Arrow points to dentary-squamosal articulation; the articular-quadrate articulation lies directly behind it (toward the midline of the skull). (*Illustration by Christine L. Huffard*)

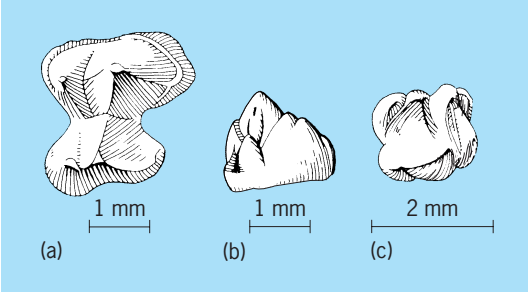

Fig. 2. Upper and lower molars of *Docodon*, a Late Jurassic docodont. (*a*) Occlusal view of an upper molar. (*b*) Internal and (*c*) occlusal views of a lower molar.

contrast to the highly derived (advanced) morphology and function of its dentition, the articulation of the lower jaws and skull of *Haldanodon* (and possibly other docodonts) exhibits a mixture of primitive and derived characters. Contact is retained between the articular bone of the lower jaw and the quadrate bone of the skull, which was the sole jaw articulation of their ancestors—the basal synapsids. The middle ear contained only one bone, the stapes (unlike the condition in other mammals, in which three bones are present). In addition to the primitive articular-quadrate articulation, a second jaw articulation, between the dentary bone of the lower jaw and the squamosal bone of the skull, had evolved. The latter articulation is present in all modern mammals in which the articular and quadrate bones form the malleus and incus of the middle ear. *See* EAR (VERTEBRATE); DENTITION.

Elements of the postcranial skeleton are known only from the Late Jurassic genus *Haldanodon*. The large size and sturdy processes for muscle attachments on the limb bones indicate the forearms were adapted for burrowing and swimming. *Haldanodon's* morphology and the depositional environment in which it was found (a coal swamp) suggest this docodont resembled modern semiaquatic desman moles and was an active swimmer that dug burrows in the banks of the ancient swamp. The nature of the sediments from which the fragmentary remains of other docodonts have been recovered suggest that they might not have been semiaquatic, but were probably small animals that scrambled through drier terrestrial environments.

The docodonts were one of several distinct lineages of basal mammals that differentiated during an evolutionary radiation that probably began in the Triassic. During the Jurassic docodonts diversified into two major groups, one limited to Asia and the other to Europe and North America. During the Cretaceous, docodonts became extinct: they were not ancestral to any other mammalian lineages. *See* ANIMAL EVOLUTION; MAMMALIA. William A. Clemens

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Dogs

All breeds of domestic dogs, the wild dogs, and related species belong to the family Canidae. Despite the various breeds of domestic dogs that are known, the scientific name for all is *Canis familiaris*. The origin of domestic dogs is obscure, but they seem to be most closely related to the wolf. The dog has been associated with humans since Neolithic times, at least 10,000 years ago, and there are indications that as early as 4000 years ago several varieties of dogs had been developed by humans. In many respects the dog is structurally primitive and shows a genetic plasticity which accounts for the many varieties. There are more than 100 breeds of domestic dogs, and their classification is based principally on their uses.

The dog is adapted to running, is digitigrade (walking on tiptoe), and has four toes on each foot with a dewclaw on the front feet of many varieties that represent the remnant of the fifth claw. The blunt claws are nonretractible and may be used for digging. The jaw is elongate, with 42 teeth, and the dental formula is I 3/3, C 1/1, Pm 4/4, M 2/3. Dogs are mature toward the end of the first year, after which the female comes into heat twice yearly, the period lasting about 2 weeks. Like many of the vertebrates, the male has an os priapi, or penis bone. Gestation is about 8 weeks, and the size of the litter varies from 2 to 16. The pups are blind and helpless at birth but grow rapidly, opening their eyes at 10 days and walking at 20. At the age of 1 month all of the milk, or deciduous, teeth have erupted. Permanent dentition appears at about 5 months.

The wild species of the family, numbering about 36 and having a wide distribution, include several wild dogs, wolves, coyotes, foxes, and jackals.

Domestic breeds. The domestic breeds are the sporting breeds, hounds, terriers, working breeds, toy breeds, and the nonsporting group.

Sporting breeds. These animals are trained and employed for retrieving or finding game. The largest group in this class is the spaniels, of which there are about 10 types varying in size from 18 lb (8 kg) in the cocker spaniel to almost 50 lb (22 kg) in the field spaniel. The spaniels have been trained to retrieve and to flush game, and they hunt both birds and fur-bearing animals. The Brittany spaniel is the only member of the group which points.

Retrievers are specially trained to locate and return game to hunters and are used most commonly in hunting waterfowl. They are usually large dogs and excellent swimmers. There are four varieties: the curly, the Chesapeake Bay, the golden retriever, and the Labrador. With the exception of the golden retriever, the Labrador is the origin of the others by crossbreeding with spaniels and setters.

Pointers and setters, used for hunting upland game birds, range ahead of the hunter, point the game until the hunter arrives, and retrieve the fowl after it has been flushed and shot. Among the varieties that are used in these pursuits are the Weimeraner, the English and Irish setters, and the German short-haired

pointer. The setter probably is derived from pointer and spaniel ancestry.

Hounds. This group includes the basset, the bloodhound, the whippet (produced by crossing a terrier with an Italian greyhound and then backbreeding to the English greyhound), the dachshund, the wolfhound, and the beagle (**Fig. 1**). The greyhound, one of the oldest breeds, is built for speed with its thin body and long legs. The bloodhounds and foxhounds (smaller, stockier dogs) are used for hunting, mostly by scent. The beagle, now more a pet than a hunter, can also follow a scent and is easier to follow on foot.

Terriers. The terriers originally were bred for hunting burrowing animals, such as the badger and fox. The Boston terrier, probably derived from the bulldog and bull terrier, is the only breed to have originated in the United States. The fox terrier, typical of the group, may have been derived from the foxhound. It was originally bred for fox hunting but is an excellent ratter. Other terriers are the Airedale, largest of the group, and the Scottish and Skye terriers.

Working breeds. Most of these breeds are large animals used as draught animals, for police work, for herding, and as guide dogs for the blind. The St. Bernard is a draught animal in the Low Countries, and is used for milk delivery and other work. Other draught breeds are the Alaskan malamute, the Eskimo, and the Samoyed.

Among the animals used as guard dogs and for police work are the Doberman, the German shepherd, and the Great Dane. All these animals are specially trained for sentry duty and guard work, but the Great Dane, one of the most powerful of dogs, is also used to herd cattle and is still employed as a draught animal in France and Belgium. The collies, Belgian sheep dogs, and English sheep dogs are outstanding sheep herders.

The bulldog is now more of a pet and house dog than a guard dog, but it is pugnacious if set on an intruder and will not release the person. The poodle, said to be the most intelligent dog, can be trained as a gun dog and was originally used for duck shooting.

Toy breeds. Some of these breeds have been known for centuries and are principally household pets which may develop an instinct for protecting the premises of the owner. They are all quite small, some being miniatures of the larger breeds. The chihuahua is the smallest, weighing as little as 1 lb (0.45 kg). Some of the more popular varieties are the Pomeranian, Pekingese, and pug.

Nonsporting breeds. Many of the dogs included in the foregoing groups are included in this category as a matter of opinion. The Dalmatian may be included in this group as well as the poodle. The chow has been raised for centuries in China, where it was used as a hunting dog as well as for its meat and fur. It is now used as a guard dog and a household pet, and is unusual in that its tongue is blue.

Wild species. There are a number of wild dogs which have never been domesticated, unlike the dingo of Australia (*Canis dingo*) that is believed to have been a domestic dog introduced into Australia during prehistoric times and then reverted to its wild state. It is the only wild carnivore in Australia. This animal has been domesticated by the natives and is used to hunt kangaroos and other game. In the wild state it occurs singly rather than in packs and is destructive to livestock.

Fig. 1. The beagle, a popular domestic breed.

The Asiatic wild dog occurs throughout Asia, Java, and Sumatra and is considered to be three distinct species by some authorities while others regard it as two subspecies of *Cuon alpinus*, which is also known as the Siberian wild dog. In India this dog is known as the dhole and has been given the name *C. dukhunensis*. It is slightly larger than the Siberian variety, which is about 3 ft (1 m) long, and lives in forested areas. The Malay variety (*C. javanicus*) is the smallest of the three, and the fur is a brighter color. These are running hunters, hunting in packs which consist of a family or a number of families that will attack any animal they encounter.

The Cape hunting dog (*Lycaon pictus*) ranges throughout the grasslands of eastern and southern Africa but has become reduced in numbers. This animal runs in packs which may contain up to 60 individuals and usually rests during the day and hunts when it is cooler. They differ from other members of the family in having four toes on the front foot. The bush dog (*Speothos venaticus*), the maned wolf (*Chrysocyon brachyurus*), and the Cordillera fox (*Cerdocyon magellanicus*), all indigenous to South America, are considered to be wild dogs despite the common names. The bush dog is about 2 ft (0.6 m) long with short legs and tail. The maned wolf, more than 4 ft (1.2 m) long with a 16-in. (40-cm) tail, resembles a giant fox and has extremely long legs. The Cordillera fox is about 3 ft (1 m) in length, excluding the tail.

Wolves. The common European wolf (*Canis lupus*) is the species that once ranged throughout the temperate forested regions of Europe, Asia, and

North America. This species has been exterminated in the British Isles and almost so in France, but they do occur in other European countries such as Italy, Spain, and the Balkans and are still plentiful in the Scandinavian countries. The coat is usually a brown-gray color, and its thickness depends upon the environment. Wolves howl; they do not bark as do dogs. The adult stands about $2^1/_2$ ft (0.75 m) tall and weighs more than 100 lb (45 kg). They inhabit hills and plains areas in addition to forested areas. The female has one litter of from three to nine cubs each year. Mating occurs at the beginning of the year, and the gestation period is about 2 months. The female usually prepares a den near some stream, lines it with vegetation, and then cares for the cubs until they are weaned at about 6 weeks. The father remains with the family and will raise the cubs if anything happens to the female. The offspring are mature at the age of 2. Wolves hunt in packs and are extremely cunning. Due to recovery efforts, the gray wolf or timber wolf, recently restricted to Alaska and the subarctic regions of Canada, has now been restored to many areas of its historic North American range—including portions of the southwest United States, the Rocky Mountains, and the western Great Lakes region.

The coyote (*Canis latrans*) is sometimes called the prairie wolf and is a close relative of the true wolf, although it is smaller than the wolves, the adult being 2 ft (0.6 m) high and weighing about 40 lb (18 kg). Coyotes have historically inhabited the prairies, open plains, and desert areas of North America; however, since the early twentieth century they have spread eastward as far as New England and are now ubiquitous throughout the contiguous United States, being found in many types of habitats. Food consists of insects, prairie dogs, and large game when they hunt in packs, but usually this animal hunts alone. It has keen vision and an excellent sense of smell and is extremely fast. The den is frequently one which has been abandoned by a fox, badger, or skunk, and it is enlarged and cleaned thoroughly. In the spring or early summer the litter of five to seven cubs is born.

Fig. 2. Red fox (*Vulpes vulpes*). (*Photo by Gerald and Buff Corsi; © 1999 California Academy of Science*)

Jackals. These animals are scavengers as well as menaces to domestic poultry. They are nocturnal in their activities and hunt in packs. Both the oriental jackal (*Canis aureus*), the most widely distributed jackal, and the black-backed jackal (*C. mesomelas*) can be easily tamed. *Canis aureus* has spread from southeastern Europe and northern Africa through Asia as far south as Burma. It prefers higher elevations in contrast to *C. mesomelas*, which is found in the grasslands of eastern and southern Africa. It is often a solitary hunter; however, several may form a pack and follow the trail of a predator such as a lion to finish the remains of a kill. The simenian jackal (*C. simensis*), a large species found in Abyssinia, is rare, and little is known of its natural history. The most common enemy of the jackal is the leopard.

Foxes. These members of the dog family have relatively short legs and long bodies, big erect ears, pointed snouts, and long bushy tails (**Fig. 2**). The red fox (*Vulpes vulpes*), found throughout Europe and the North America as well as parts of Asia, North Africa, and Australia, actually has undergone many color phases and mutations and includes other varieties such as the silver fox and the cross fox. *Vulpes vulpes* stands a few inches over 1 ft (0.3 m) and weighs from 15 to 20 lb (6.5 to 9 kg). Scent glands are present in the anal region, which account for the characteristic odor of these animals.

Foxes may use an abandoned den of another animal, dig their own den in a protected location, or live aboveground in a secluded hollow under rocks or in tree hollows. These latter abodes are referred to as kennels. The fox is active nocturnally and preys upon small rodents, birds, and rabbits; if it lives near a stream, it also eats crayfish and fish.

The male may hunt alone or may pair with the female, known as a vixen, to chase the quarry. Mating occurs in the winter, and the gestation period is 8 weeks. The litter varies from three to eight cubs, which are born blind. The cubs are weaned at the age of 2 months, and the male assists in feeding them thereafter. They remain with the parents until autumn, learning to hunt and fend for themselves, and mature for the following year. Near farm areas they may be a nuisance to poultry breeders since they may raid coops.

Most species of foxes provide good fur, and with the destruction of natural wildlife environments and the increased demand for fur, many fox farms were started in Canada and the Scandinavian countries as well as in the northern areas of the United States. These animals breed well in captivity. *See* CARNIVORA; MAMMALIA; SCENT GLAND. Charles B. Curtin

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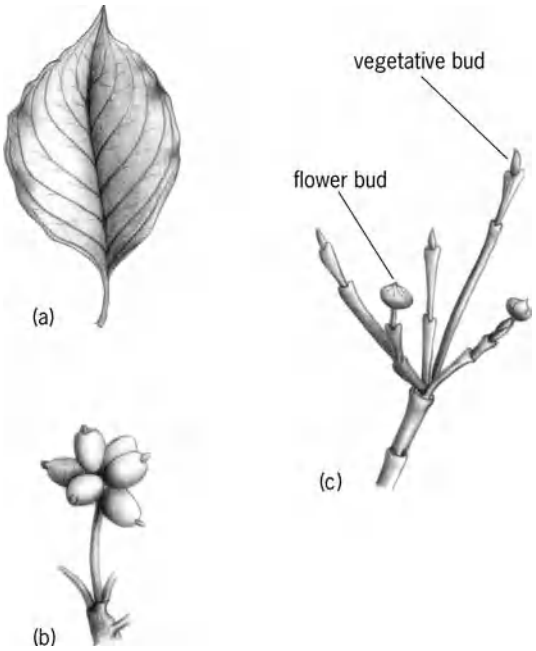
Dogwood

A tree, *Cornus florida*, also known as flowering dogwood, which may reach a height of 40 ft (12 m) and is found in the eastern half of the United States and in southern Ontario, Canada. It has opposite, simple, deciduous leaves with entire margins (see **illus.**).



Flowering dogwood (*Cornus florida*), with characteristic (*a*) leaf, (*b*) fruit cluster, and (*c*) branch.

When this tree is in full flower, the four large, white, notched bracts or petallike growths surrounding the small head of flowers give an ornamental effect that is unequaled by any native tree. Pink, rose, and cream-colored varieties are commonly planted. The tree is tolerant of shade and at blossoming time, usually in early May, the patches of white reveal its presence, even in dense woods. The wood is very hard and is used for roller skates, carpenters' planes, and other articles in which hardness is desired. As a shade tree, it is especially desirable for the modern ranch-type house where small size is appropriate. The Pacific dogwood (*C. nuttalli*), which grows in Idaho and from southwestern British Columbia to southern California, is similar to the eastern dogwood, but has rounded bracts. The Japanese dogwood (*C. kousa*) is a similar small tree with pointed bracts and blooms in June. Other shrubby species of dogwood are used as ornamentals. *See* CORNALES; FOREST AND FORESTRY; TREE.

Arthur H. Graves; Kenneth P. Davis

Dolerite

A fine-textured, dark-gray to black igneous rock composed mostly of plagioclase feldspar (labradorite) and pyroxene and exhibiting ophitic texture. It is commonly used for crushed stone. Its resistance to weathering and its general appearance make it a first-class material for monuments. *See* STONE AND STONE PRODUCTS.

The most diagnostic feature is the ophitic texture, in which small rectangular plagioclase crystals are enclosed or partially wrapped by large crystals of pyroxene. As the quantity of pyroxene decreases, the mineral becomes more interstitial to feldspar. The rock is closely allied chemically and mineralogically with basalt and gabbro. As grain size increases, the rock passes into gabbro; as it decreases, diabase passes into basalt. *See* BASALT; GABBRO.

Diabase forms by relatively rapid crystallization of basaltic magma (rock melt). It is a common and extremely widespread rock type. It forms dikes, sills, sheets, and other small intrusive bodies. The Palisades of the Hudson, near New York City, are formed of a thick horizontal sheet of diabase. In the lower part of this sheet is a layer rich in the mineral olivine. This concentration is attributed by some investigators to settling of heavy olivine crystals through the molten diabase and by others to movement of early crystals away from the walls of the passageway along which the melt flowed upward from depth, before it spread horizontally to form the sill. *See* MAGMA.

As defined, diabase is equivalent to the British term dolerite. The British term diabase is an altered diabase in the sense defined here. *See* BASALT; GABBRO; IGNEOUS ROCKS.

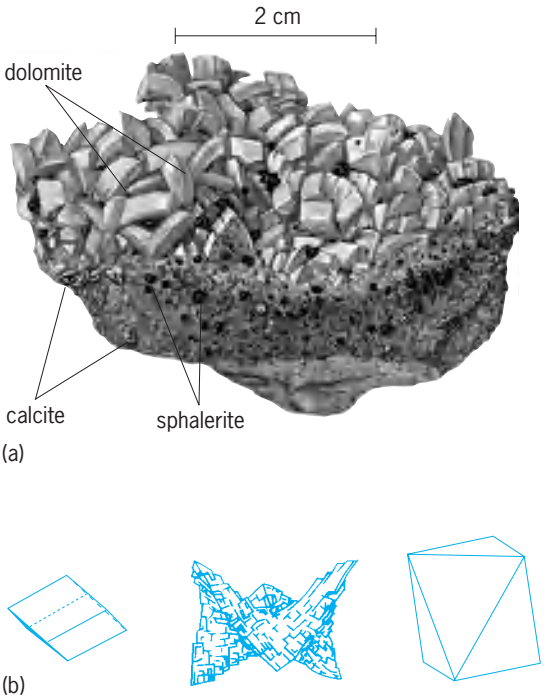
Carleton A. Chapman

Dolomite

The carbonate mineral $CaMg(CO_3)_2$. Often small amounts of iron, manganese, or excess calcium replace some of the magnesium; cobalt, zinc, lead, and barium are more rarely found. Dolomite merges into ankerite with increasing substitution of iron for magnesium, and into kutnahorite with increasing manganese substitution. *See* ANKERITE.

Dolomite has hexagonal (rhombohedral) symmetry and a structure similar to that of calcite, but with alternate layers of calcium ions being completely replaced by magnesium. Thus each layer of CO_3^{2-} ions lies between a layer consisting entirely of Ca^{2+} ions and one consisting entirely of Mg^{2+} ions. This ordered arrangement of the cations distinguishes the structure of dolomite from that of a randomly substituted calcite. *See* CALCITE.

Dolomite is a very common mineral, occurring in a variety of geologic settings. It is often found in ultrabasic igneous rocks, notably in carbonatites and serpentinites, in metamorphosed carbonate sediments, where it may recrystallize to form dolomite marbles, and in hydrothermal veins. The primary occurrence of dolomite is in sedimentary deposits, where it constitutes the major component of dolomite rock



Dolomite. (*a*) Crystals with calcite and sphalerite on limestone, which were found in Joplin, Missouri (*specimen from Department of Geology, Bryn Mawr College*). (*b*) Crystal habits (*after C. S. Hurlbut, Jr., Dana's Manual of Mineralogy, 17th ed., John Wiley and Sons, 1959*).

and is often present in limestones (see **illus.**). *See* DOLOMITE ROCK; LIMESTONE; SEDIMENTARY ROCKS.

Dolomite is normally white or colorless with a specific gravity of 2.9 and a hardness of 3.5–4 on Mohs scale. It can be distinguished from calcite by its extremely slow reaction with cold dilute acid. *See* CARBONATE MINERALS.

Alan M. Gaines

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Dolomite rock

Sedimentary rock containing more than 50% by weight of the mineral dolomite [$CaMg(CO_3)_2$]. The term dolostone is used synonymously. A range of the Alps in northeastern Italy is named the Dolomites because it is composed predominantly of dolomite rock. *See* DOLOMITE.

Characteristic properties. Since dolomites usually form by replacement of preexisting limestones, it is often possible to see relict grains, fossils, and sedimentary structures preserved in dolomite rocks. More often, however, original textures are obliterated. *See* LIMESTONE.

There is a very strong tendency for dolomite to form euhedral rhombs. This is useful for distinguishing dolomite from the usually anhedral calcite. Dolomite rhombs vary in size from micrometers (on edge) to millimeters. These rhombs often contain cloudy, inclusion-filled cores and clearer rims. Trace-element zonation is also common, as can be seen in stained thin sections or by cathodoluminescence. *See* PETROGRAPHY.

Field differentiation of dolomite from calcite is most easily accomplished by application of dilute hydrochloric acid. Calcite is strongly effervescent in acid, whereas dolomite is usually very weakly effervescent unless scratched or ground into a fine powder. Dolomite also often weathers to a brown color on outcrop because of the common substitution of iron for magnesium in the dolomite structure. X-ray diffraction analysis is the most reliable method for the differentiation of dolomite from calcite, and also the best method for characterization of the degree of ordering of a particular dolomite sample. *See* X-RAY DIFFRACTION.

Dolomite rocks are predominantly monomineralic. The most common noncarbonate minerals are quartz (either authigenic or detrital), clay minerals, pyrite, and glauconite. Evaporite minerals or their replacements are also common.

Occurrence and origin. Geologists have been unable to decipher the exact conditions of dolomite formation. This so-called dolomite problem revolves around several questions relating to the stoichiometry of the reaction in which dolomite is formed, the fact that dolomite is not common in young marine sediments, and the type of geological setting in which ancient dolomites formed.

If dolomite formed in nature as a primary or direct precipitate from constituent ions dissolved in aqueous solutions, dolomitization could be written as reaction (1). Alternatively, if dolomite formed as a

$$Ca^{2+} + Mg^{2+} + 2CO_3{}^{2-} \longrightarrow CaMg(CO_3)_2 \qquad (1)$$

secondary mineral by replacement of precursor aragonite or calcite (two different minerals with the same chemical composition: $CaCO_3$), the dolomitization reaction could be represented as (2). A third, intermediate option is given in reaction (3). There is no

$$2CaCO_3 + Mg^{2+} \longrightarrow CaMg(CO_3)_2 + Ca^{2+} \qquad (2)$$

$$CaCO_3 + Mg^{2+} + CO_3{}^{2-} \longrightarrow CaMg(CO_3)_2 \qquad (3)$$

evidence for dolomite forming as a primary precipitate within ocean (or lake) water by reaction (1). Hydrothermal syntheses of dolomite at temperatures of 100°C (212°F) and greater have proven that dolomite formation can take place by the replacement of calcium carbonate [reaction (2)]. Reaction (3) is almost certainly the most important in nature: most of the calcium in dolomite originates in a solid calcium carbonate precursor, but no extra calcium is released to solution. *See* ARAGONITE; CALCITE.

Examination of reaction (2) indicates that the higher the Mg^{2+}/Ca^{2+} ratio in solution, the greater the tendency for dolomite to replace calcium carbonate. If the ratio is too high, however, magnesite ($MgCO_3$) may replace dolomite [reaction (4)].

$$CaMg(CO_3)_2 + Mg^{2+} \longrightarrow 2MgCO_3 + Ca^{2+} \qquad (4)$$

The calculated range of stability of dolomite over

calcite and magnesite (at 25°C or 77°F) is between Mg^{2+}/Ca^{2+} ratios of about 0.7 and 5.5, respectively. Since many natural waters, including seawater, have Mg^{2+}/Ca^{2+} ratios within this range, dolomite should be the stable carbonate mineral in nature. Surprisingly, dolomite is very rare in young carbonate sediments, which are instead dominated by aragonite and calcite.

The relative propositions of dolomite and calcite have changed through geologic time. Dolomitic rocks are dominant in the Precambrian and early Paleozoic, whereas calcitic rocks become dominant in the late Mesozoic and continue to dominate through the present. Several alternative hypotheses have been offered to explain this. Perhaps the composition of Precambrian oceans was different from that of present-day oceans (either in their dissolved CO_2 content, their Mg^{2+}/Ca^{2+} ratio, or even their dissolved sulfate content). Perhaps just the increased chance of exposure of calcitic precursor rocks to dolomitizing solutions ensured that older rocks would be more dolomitic. Perhaps the mineralogy of potential dolomitic precursors has changed through geologic time. Aragonite dolomitizes much more rapidly than calcite; hence a shift in the dominant mineralogy of carbonate sediments from aragonite to calcite in the Mesozoic could favor the dominance of calcitic rocks over dolomitic ones.

In a few geological settings, dolomite is forming within the sediments at the present time. One general setting includes a wide variety of supratidal ponds, lagoons, tidal flats, and sabkhas such as the Solar Lake in Israel, the Coorong Lagoon in Australia, the bank tops of Andros Island in the Bahamas, and the Sabkha of Abu Dhabi. Each of these environments is hot, more saline than seawater, and rich in organic matter. A second general setting of modern dolomite formation is in continental margin sediments underlying productive coastal oceanic upwelling zones. Examples include the Peru–Chile margin, the southern California borderlands, the Gulf of California, and the Walvis Ridge west of Namibia (southwestern Africa). This second set of geological environments is characterized by lower water temperature and normal marine salinity. Like the first set, these environments are rich in organic matter. The association of dolomite formation with abundant organic carbon seems to be the one common thread linking each dolomite occurrence and possibly controlling dolomite precipitation.

Previously, researchers had focused a great deal of attention on so-called dolomitization models, which were largely concerned with how to get sufficient reactants to the site of dolomite formation. Significant amounts of dolomite formation require efficient removal of Mg^{2+} from very large volumes of seawater. A successful model therefore requires a mechanism to move large amounts of seawater over buried calcium carbonate sediments. For some reason, ordinary seawater does not permit the formation of dolomite. It is now believed that the dolomitization reaction is very slow, but can be accelerated by the removal from seawater of dissolved sulfate (which occurs during the postburial breakdown of sedimentary organic matter) and by high temperatures.

Economic implication. Dolomites are extremely important oil reservoir rocks. This is partly a result of the high porosity of many dolomite rocks and partly the result of the association between dolomite formation and the presence of sedimentary organic matter.

Dolomites are also the main host rocks for lead and zinc ore deposits. Rocks of this type are known as Mississippi Valley ore deposits. Neither the origin of the ore-forming solutions nor the association between the lead-zinc ores and the dolomite rocks has been satisfactorily explained. *See* PETROLEUM GEOLOGY; SEDIMENTARY ROCKS.

Paul A. Baker

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Domain (electricity and magnetism)

A region in a solid within which elementary atomic or molecular magnetic or electric moments are uniformly aligned.

Ferromagnetic domains are regions of parallel-aligned magnetic moments. Each domain may be thought of as a tiny magnet pointing in a certain direction. The relatively thin boundary region between two domains is called a domain wall. Within a wall the magnetic moments rotate from the direction of one of the domains to the direction in the adjacent domain.

A ferromagnet generally consists of a large number of domains. For example, a sample of pure iron at room temperature contains many domains whose directions are distributed randomly, making the sample appear to be unmagnetized as a whole. Iron is called magnetically soft since the domain walls move easily if a magnetic field is applied. In a magnetically hard or permanent magnet material (for example, alnico or neodymium-iron-boron) a net macroscopic magnetization is introduced by exposure to a large external magnetic field, but thereafter domain walls are difficult to either form or move, and the material retains its overall magnetization.

Antiferromagnetic domains are regions of antiparallel-aligned magnetic moments. They are associated with the presence of grain boundaries, twinning, and other crystal inhomogeneities. *See* ANTIFERROMAGNETISM; TWINNING (CRYSTALLOGRAPHY).

Ferroelectric domains are electrical analogs of ferromagnetic domains. *See* FERROELECTRICS; FERROMAGNETISM; MAGNETIZATION.

J. F. Herbst

Domestication (anthropology)

The adaptation of an animal or plant through breeding in captivity to a life intimately associated with and advantageous to humans. Domesticated plants and animals differ genetically and morphologically from their wild counterparts, often in ways that make them more useful to humans. For example, wild sheep are hairy but domesticated sheep are woolly. Some of the earliest domesticated animals include sheep, goats, cattle, pigs, and horses, which were domesticated in the Old World, and llamas and alpacas, which were domesticated in the Americas. Dogs are known from both the Eastern and Western hemispheres. Important early domesticated plants include cereals—wheat, barley, rice, sorghum, and millet in the Old World and maize in the Americas—and tubers such as yams and taro in the Old World and potatoes in the Americas. Other important domesticated plants include pulses, olives, cotton, flax, and fruits. Domesticated plants and animals provide humans with a variety of useful products, including food and fibers, and serve important functions, such as animal traction (that is, the use of animals to pull carts and plows).

Agricultural revolution. The beginnings of plant and animal domestication is sometimes referred to as the agricultural or Neolithic revolution. (The term "Neolithic" means new stone age and refers to the appearance of ground and polished stone axes which are commonly found on early agricultural sites.) Analogous to the industrial revolution, the agricultural revolution represented a change in subsistence technology that affected many aspects of human societies. The beginning of agriculture and animal husbandry was a significant turning point in human history. Until the end of the Ice Age, about 11,500 years ago, all humans throughout the world lived by foraging, that is, hunting animals and gathering plants. At the time of Columbus's voyages, about 500 years ago, it is estimated that about 15% of the world's population lived by hunting and gathering. Today almost no hunters and gatherers remain, and the bulk of the world's food supply comes from agricultural produce.

The study of the beginnings of agriculture has intrigued archeologists and anthropologists for over a century because agriculture provided the economic basis for the development of complex urban societies. Maize was a staple crop throughout the Americas in later pre-Columbian times, and cereals such as wheat, barley, and rice were critical to the development of complex societies in the Eastern Hemisphere. Farming provides one major advantage over hunting and gathering—a higher yield per acre. The surplus food provided by early farmers could be used to support craftsmen and other nonfood producers and to finance the construction of monumental public works such as temples and pyramids.

The reasons why human societies began to cultivate crops and herd animals are not well understood; many possible explanations for the origins of farming have been put forward. Since the early part of the twentieth century, some archeologists have suggested that the climatic changes at the end of the Ice Age either encouraged or forced early human populations to experiment with agriculture because they affected the distribution of wild plants and animals. Other scholars have suggested that population growth may have led to the adoption of farming. Still others have focused on social factors, suggesting that the social demands of competitive feasting and exchange may have encouraged innovations that led to higher food yields. Today most archeologists reject a monocausal explanation for the beginnings of farming. Environmental, demographic, and social factors undoubtedly played very different roles in the beginnings of farming in various parts of the world. *See* AGRICULTURAL SCIENCE (ANIMAL); AGRICULTURAL SCIENCE (PLANT); AGRICULTURE; NEOLITHIC.

Methods of study. Studies of plant and animal domestication are most often carried out by zooarcheologists and paleoethnobotanists, archeologists with specialized expertise in animal bones and plant remains. In order to study the origins and spread of plant and animal domestication, archeologists rely on plant and animals remains that are recovered from archeological sites. Animal bones are some of the most common finds on archeological excavations. They are generally recovered using fine screens so that the bones of small animals are not lost. Evidence for animal domestication can include the discovery of an animal outside its wild range, morphological changes such as size reduction, and changes to the demographic structure of the animal herd. Plant remains are generally recovered using the flotation process, in which archeological soils are poured into moving water. Light organic materials, such as charred seeds, are held in suspension momentarily so that they can be poured off into a fine-mesh sieve. In cereal crops, plant domestication can lead to changes in the rachis, the axis of the plant that holds the grains. Wild plants have a brittle rachis, allowing the plants to reseed themselves when the grains ripen, whereas domestication can produce a nonbrittle rachis. *See* ARCHEOLOGY; PALEOBOTANY.

Centers of domestication in the Middle East. Studies of agricultural origins have focused on the Near and Middle East because the ancient Near East was home to some of the world's earliest complex societies, including the Sumerians and Akkadians. In addition, archeological research has shown that the Near East is home to some of the world's earliest farmers. The Natufian culture of the southern Levant, the region along the east coast of the Mediterranean, is seen as pivotal in our understanding of the transition from foraging to farming.

Environmental studies have shown that climatic amelioration at the end of the Ice Age, beginning about 14,000 years ago, led to an expansion of Mediterranean vegetation throughout the region. The Mediterranean vegetation zone is characterized as an oak-pistachio parkland and includes the wild ancestors of both emmer wheat (*Triticum dicoccoides*) and barley (*Hordeum spontaneum*)

[**illus.** *a* and *c*]. Natufian foragers appear to have collected wild cereals and other plant foods such as acorns and to have hunted gazelles and other large mammals between about 14,000 and 11,500 years ago. Although the Natufians made their living by hunting and gathering, they did keep domestic dogs (*Canis familiaris*). *See* DOGS; WHEAT.

In parts of the southern Levant, at sites such as Mallaha/Eynan in northern Israel and Wadi Hammeh 27 in Jordan, Natufians inhabited small villages on a year-round basis. Colder and drier conditions toward the end of the Ice Age, about 12,000 years ago, may have disrupted the Natufian way of life, especially in the drier, more marginal parts of the southern Levant. Some archeologists have suggested that these environmental changes may have encouraged experimentation with cereal cultivation. *See* ICE AGE.

Plant cultivation. Shortly after the end of the Ice Age, foragers throughout the southern Levant appear to have started experimenting with cereal cultivation. Sites such as Netiv Hagdud in the Jordan Valley have provided archeological evidence for the earliest stages of plant cultivation. The cereal grains recovered from this site, dated from about 9000 B.C., are morphologically wild. They have a brittle rachis, which allows the barley to reseed itself. The case for cultivation is based on the high proportion of barley grains recovered at Netiv Hagdud (about 90% of the grasses at the site). Possible domestic barley grains appear at the site of Tell Aswad in Syria, dated to about 7700 B.C., but the earliest well-dated barley was recovered from Ain Ghazal in Jordan and dates to about 7000–6500 B.C. Domesticated emmer wheat was also widespread in the Near East by 7300 B.C. Barley and emmer wheat were among the earliest domesticated plants (illus. *a* and *c*). Deoxyribonucleic acid (DNA) and archeological evidence suggest that a second species of wheat, einkorn (*T. monococcum*), was initially domesticated in Anatolia (illus. *b*). During the early Neolithic period, shortly after the domestication of cereals, pulses such as lentils and bitter vetch were also domesticated in the Middle East.

The archeological evidence suggests that plant domestication preceded animal domestication in the Middle East. Early sites such as Netiv Hagdud provide evidence for early plant cultivation combined with the hunting of wild animals. *See* BARLEY; CEREAL.

Animal domestication. The earliest evidence for animal domestication in the Near East appears in the early eighth millennium B.C. Archeological evidence for domestic goats first appears at the site of Ganj Dareh in Iran (illus. *d*). In general, early herders treated male and female goats somewhat differently. Since females are the productive members of the herd, providing both milk and offspring, female animals typically survive into adulthood. Surplus male animals are often slaughtered as juveniles. The kill patterns for early Neolithic goats from Ganj Dareh reflect this differential treatment of males and females. During the seventh millennium B.C., goats from a number of sites throughout the Near East show morphological changes, such as twisting of the horn cores (the bony portion of the skull that lies underneath the horn), that are hallmarks of animal domestication. Other early Near Eastern animal domesticates include sheep, followed by cattle and pigs (illus. *e*). *See* GOAT PRODUCTION.

Primary Near Eastern domesticates. (*a*) Emmer wheat, *Triticum dicoccum* (wild from *T. dicoccoides*). (*b*) Einkorn wheat, *T. monococcum* (wild form *T. boeoticum*). (*c*) Barley, *Hordeum vulgaris* (wild form *H. spontaneum*). (*d*) Goats, *Capra hircus* (wild form *C. aegagrus*, the Bezoar goat). (*e*) Sheep, *Ovis aries* (wild form *O. orientalis*, the Asiatic mouflon). (*Reprinted with permission from P. J. Crabtree and D. V. Campana, Archaeology and Prehistory, McGraw-Hill, 2001*)

Other early centers of domestication. While the Near East appears to be home to the earliest farmers in the Eastern Hemisphere, a number of other early centers of plant and animal domestication have been identified. Rice was initially domesticated in southeast Asia, while broomcorn millet (*Panicum miliaceum*) was first domesticated in north China. Recent archeological research has suggested that New Guinea farmers may have begun to plant taro and bananas as early as 5000 B.C. A variety of plants, including sorghum, pearl millet, yams, and watermelon, were initially domesticated in sub-Saharan west Africa, possibly beginning about 4500 B.C. In addition, recent DNA studies have suggested that Africa just south of the Sahara may have been an independent center of cattle domestication as early as 10,000 years ago. Humped cattle (*Bos indicus*) were first domesticated in South Asia. Later Eurasian domesticates include horses, which were probably domesticated in the Russian/Ukrainian steppe

zone about 3500 B.C., and camels, which were domesticated in both eastern Iran (Bactrian, or two-hump, camels) and Arabia (dromedary, or one-hump, camels) about 2500 B.C. The domestication of the olive and the vine transformed Mediterranean agriculture, since olives and grapes can be grown on land that is unsuited to cereal crops. Biochemical analyses have identified wine residues from an early pottery vessel from Iran dated to 5400 B.C. Widespread finds of transport amphorae, large ceramic vessels used to transport liquids, indicate that wine and olive oil were widely traded throughout the ancient world. *See* BEEF CATTLE PRODUCTION; CAMEL; OLIVE.

Centers of domestication in the Americas. Archeologists working in the Americas have been especially interested in studying the domestication of maize, as this crop was a staple food for the Aztec, Inca, Maya, and other complex societies of the New World.

Mesoamerica. During the 1960s, the Tehuacán Valley project was designed to study the origins of plant domestication in highland southern Mexico. The initial results suggested that domesticated maize first appeared in that region about 3500 B.C. However, redating of the early domesticated corn cobs from the Tehuacán Valley using the accelerator mass spectrometer (AMS) process has shown that these cobs are nearly 1000 years younger than was originally thought. Today the earliest known maize cobs come from the site of Guilá Naquitz in Oaxaca and date to about 4300 B.C. This cave site has also produced the earliest known domesticated squash (*Cucurbita pepo*) from the region. This squash is dated to 8000–6000 B.C., making it several millennia older than the earliest domesticated maize.

Recent archeological research on maize domestication has focused on the Balsas River Valley in the Pacific highlands of southern Mexico, as DNA research has shown that the teosinte (*Zea mays* ssp. *parviglumis*, the wild ancestor of corn) from this region is most closely related to domesticated corn. Since the corn cobs from Guilá Naquitz are unquestionably domesticated, the initial cultivation of maize must have begun appreciably earlier than 4300 B.C. *See* CORN.

Other plants that were domesticated in Mesoamerica, the geographical region that includes both southern Mexico and Central America, include beans, peppers, and avocados. While a wide range of plants were cultivated in Mesoamerica, domesticated animals played only a limited role in the pre-Columbian economies of Central America. The most important animal domesticates were the dog and the turkey.

Highland South America. Highland South America was a separate center of plant and animal domestication in prehistoric times. Plants that were first domesticated in this region include quinoa (*Chenopodium quinoa*), a small-seeded cereallike plant, and potatoes. Potatoes are an especially useful crop because they can grow on relatively poor soils. Both the llama (*Lama glama*) and the alpaca (*Lama pacos*) were initially domesticated in the Andean region. The llama is a multipurpose animal that was used for meat and for bearing burdens. The alpaca was raised primarily for its fine wool. In later prehistoric times, alpaca wool was combined with cotton, which was grown in the arid regions along the Peruvian coast, to produce fine textiles. *See* ALPACA; LLAMA.

Eastern South America and eastern North America. Two other centers of plant domestication have been identified in the Americas. Manioc (*Manihot escuelenta*), a starchy root crop, was initially domesticated in the tropical rainforest region of eastern South America. A number of small-seeded plants were domesticated in eastern North America, including sunflowers, sumpweed (*Iva annua*), and goosefoot (*Chenopodium berlandieri*). Native American farmers began to domesticate these seeds around 2500 B.C., two millennia before maize was introduced to the eastern United States.

Conclusion. The beginning of farming was a true revolution in human subsistence and ways of life. Most of the foods we eat today would have been unknown to hunters and gatherers at the end of the Ice Age. Archeological research conducted since world War II has shown that plants and animals were domesticated in many different regions of the Eastern and Western hemispheres. Ongoing archeological research will continue to shed light on the agricultural revolution and its consequences. Pam J. Crabtree

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Dominance

The expression of a trait in both the homozygous and the heterozygous condition. In experiments with the garden pea, the Austrian botanist Gregor Mendel crossed plants from true-breeding strains containing contrasting sets of characters. For seed shape, round and wrinkled strains were used. When plants with round seeds were crossed to plants with wrinkled seeds (P_1 generation), all offspring had round seeds. When the offspring (F_1 generation) were self-crossed, 5474 of the resulting F_2 offspring were round and 1850 were wrinkled. Thus, the round trait is expressed in both the F_1 and F_2 generations, while the wrinkled trait is not expressed in the F_1 but is reexpressed in the F_2 in about one-fourth of the offspring. In reporting these results in a paper published in 1866, Mendel called the trait which is expressed in the F_1 generation a dominant trait, while the trait which is unexpressed in the F_1 but reappears

in the F_2 generation was called a recessive trait. *See* MENDELISM.

Traits such as round or wrinkled are visible expressions of genes. This visible expression of a gene is known as the phenotype, while the genetic constitution of an individual is known as its genotype. The alternate forms of a single gene such as round or wrinkled seed shape are known as alleles. In the P_1 round plants, both alleles are identical (since the plant is true-breeding), and the individual is said to be homozygous for this trait. The F_1 round plants are not true-breeding, since they give rise to both round and wrinkled offspring, and are said to be heterozygous. In this case, then, the round allele is dominant to the wrinkled, since it is expressed in both the homozygous and heterozygous condition. Dominance is not an inherent property of a gene or an allele, but instead is a term used to describe the relationship between phenotype and genotype. *See* ALLELE; GENE ACTION.

Incomplete dominance. The production of phenotypes which are intermediate between those of the parents is an example of partial or incomplete dominance. For example, in *Mirabilis,* the four-o'clock, plants with red flowers crossed to plants with white flowers produce pink offspring. Since in this case only enough red pigment is made to give a pink color, dominance is said to be partial or incomplete. Because of this lack of dominance, the phenotype of the heterozygote can be distinguished from that of either of the homozygous parents. In other words, the phenotypic ratio is equal to the genotypic ratio in both the F_1 and F_2 generations.

The phenomenon of incomplete dominance which results in a clear-cut intermediate phenotype is relatively rare. However, even in cases where dominance appears to be complete, there is often evidence for intermediate gene expression. For example, in the human biochemical disorder phenylketonuria, homozygous recessive individuals are severely affected, while heterozygotes are phenotypically normal. Homozygous recessive individuals have little or no activity of the enzyme phenylalanine hydroxylase. Heterozygous individuals have only 50% of the normal level of enzyme activity, yet are unaffected by the disease. *See* PHENYLKETONURIA.

Codominance. The separate and distinct expression of both alleles of a gene is an example of codominance. This is a situation unlike that of incomplete dominance or complete dominance. In humans, the MN blood group is characterized by the presence of molecules called glycoproteins on the surface of red blood cells. These molecules or antigens contribute to the immunological identity of an individual. In the MN blood system, persons belong to blood groups M, MN, or N. These phenotypes are produced by two alleles, M and N, each of which controls the synthesis of a variant glycoprotein. In the heterozygote MN, there is separate and complete expression of each allele. This is in contrast to incomplete dominance, where there is an intermediate or blending effect in heterozygotes. Codominance usually results in the production of gene products of both alleles. *See* BLOOD GROUPS.

Overdominance. Individuals in which the phenotype of the heterozygote is more extreme than in either of the parents are said to exhibit overdominance. In *Drosophila*, for example, the gene for white eye (*w*) when in the heterozygous condition causes an increase in the production of certain fluorescent pigments resulting in levels that are greater than either the wild type or the homozygous recessive genotypes. Examples of overdominance at the level of enzymes are also available. In the land snail *Cepaea hortilis*, homozygotes for alleles of the enzyme isocitrate dehydrogenase display four to five times less activity than do heterozygotes. In corn, alleles for the enzyme alcohol dehydrogenase when homozygous can produce forms that are stable and have low activity, or forms that are unstable with high activity. Heterozygotes between an unstable and a stable inactive form exhibit both stability and high activity. The concept of overdominance is also important in understanding the genetic structure of populations and is usually related to characteristics associated with fitness, such as size and viability.

Heterosis. The production of superior hybrid offspring by crossing two different strains of an organism is known as heterosis. The hybrid superiority may take the form of increased resistance to disease or greater yield in grain production. Heterosis in crop plants, such as corn, has been intensively studied, and the use of hybrid corn in the United States has more than doubled the average yield of bushels per acre. This has been accomplished by first establishing many highly inbred lines and then crossing them in all combinations to yield hybrids. The resulting hybrids are screened for increased productivity. Those with the highest yield are then used for crop production. *See* BREEDING (PLANT).

The mechanism which results in heterosis has been widely debated but is still unknown. Two theories have been advanced to explain the origin of hybrid superiority. One is based on the concept of overdominance. It is clear that heterozygotes for single gene differences may be superior to either of the homozygotes. At the molecular level, this overdominance may be caused by the production of metabolic products or enzymes which have combined the advantages of substances produced by either homozygote. Using this as a background, some researchers have argued that hybrids heterozygous at a large number of loci are able to thrive in environmental conditions too extreme for either homozygote. In other words, highly heterozygous genotypes are superior to homozygous ones, and this idea is called the overdominance theory of heterosis. This theory was first proposed in the early 1900s and is based on work with corn and other plants.

The alternate explanation for heterosis, known as the dominance hypothesis, is based on the idea that most recessive alleles are deleterious, while the dominant alleles of that gene are not. In this view, inbreeding produces a situation in which individuals become highly homozygous for alleles which include

many deleterious recessive traits. When these highly inbred lines are crossed, these deleterious alleles are masked by dominant alleles contributed by the opposite strain. In this way, then, the hybrid becomes superior to either homozygote.

It is difficult to determine which hypothesis is correct, and it may be that either or both mechanisms may operate in individual circumstances. In some systems, dominance can explain heterosis, while in other cases, overdominant heterozygotes are superior to either parental strain. *See* HETEROSIS.

Dominance hierarchies. In some cases, a population may carry a large number of alleles for a given gene. Although an individual carries only two alleles, within other members of a population many alternate forms of the same gene may exist. In systems involving multiple alleles, there is often a hierarchy or ordering of dominance relationships. The phenotypes for color in the peppered moth, *Biston betularia*, are a classic example of a dominance hierarchy. In *Biston*, there are three color forms, with a dominance ordering of black, gray, and white. In the land snail *Cepaea*, shell color involves a dominance hierarchy for a large number of alleles in the following order: brown, dark pink, pink, pale pink, dark yellow, yellow, pale yellow, and white. In such a series of multiple alleles which generate distinct phenotypes, each is dominant to all others below it in the hierarchy and is recessive to all phenotypes above it.

Dominance and haploids. Since dominance depends on the relationship between two alleles, it is a phenomenon usually limited to diploid organisms, or to those that have a diploid phase in their life cycle. In haploid organisms such as bacteria, however, a transitory diploid condition can be achieved. Genetic recombination in bacteria through events such as transformation, transduction, or conjugation involve transfer of all, or part of, a bacterial chromosome from one cell to another. The recipient cell is thus diploid for the transferred genes. Such cells, called merozygotes or partial zygotes, do not remain diploid. Genetic recombination takes place, and some of the transferred genetic material is incorporated into the host chromosome. All nonintegrated sequences are then lost from the cell, and a haploid recombinant cell results. *See* BACTERIAL GENETICS; CHROMOSOME; RECOMBINATION (GENETICS).

Partial diploids have been used in bacteria to explore the mechanisms of gene regulation. In the *lac* operon of *Escherichia coli,* the operator, which is a DNA sequence that regulates the transcription of three adjacent genes, can effectively regulate only the genes which are adjacent to it in the same chromosome. When a functional operator (without adjacent structural genes) is introduced into a cell which has a nonfunctional operator adjacent to three wild-type structural genes, the transferred operator is not able to exert control over the host's structural genes. As a consequence, the operator is said to be cis-dominant, because it can regulate only the genes which are adjacent to it on the same DNA molecule. *See* OPERON.

Dominant mutations. In general, mutations of a gene can cause a phenotypic change in two ways: the gene may be deleted or have a reduced or absent function (loss of function); or the gene product may function in an abnormal way or take on a new function (gain of function).

Most often, loss-of-function mutations result in a recessive phenotype, because in heterozygotes a 50% reduction in the amount of a gene product has no effect. However, in cases where half the normal level of gene product is not sufficient, a loss-of-function mutation in one allele can produce a dominant phenotype, called a dominant dosage effect. Several dominantly inherited human disorders, including Marfan syndrome, aniridia, and hypercholesterolemia, are examples of such mutations. For example, in hypercholesterolemia heterozygotes, half of the cell surface receptors that remove cholesterol from the bloodstream are defective, raising blood cholesterol levels, which in turn leads to atherosclerosis and heart disease.

Most dominant gain-of-function mutations produce a new phenotype when an allele functions abnormally (wrong time, wrong cell type, wrong amount). Very few dominant mutations produce a phenotype by taking on a new function. Human disorders classified as dominant gain-of-function mutations include McCune-Albright syndrome, where a mutant receptor is switched on at the wrong time, resulting in precocious puberty; Charcot-Marie-Tooth disease, caused by overexpression of a gene product (wrong amount); and Huntington's disease, one of the few cases where a mutant allele may have a new function.

Some gain-of-function mutations produce an abnormal phenotype when the mutant allele is not functional, and also prevent the product of the normal allele from functioning. This situation is called a dominant-negative effect. These effects are usually seen when the gene product is a protein that is composed of subunits and that functions in the form of a dimer or a multimer. Some forms of osteogenesis imperfecta, a dominantly inherited connective tissue disorder in humans, are caused by abnormalities in multimeric proteins.

Theories. The question of the origin and evolution of dominance was first examined by R. A. Fisher in 1928. He reasoned that most mutations are recessive and deleterious. To bring the phenotype of the heterozygote to a state which is identical to that of the dominant homozygote, he proposed that modifiers at other genetic loci are selected for, and that these modifiers act on the expression of the heterozygous phenotype, making it resemble that of the wild-type or homozygous dominant individual. This proposal was criticized by J. B. S. Haldane and others because of the degree of selection that would be required.

A physiological explanation of dominance was put forward by S. Wright in 1934. This theory attempted to account for dominance at a biochemical level by considering the kinetics of enzyme reactions and the existence of metabolic pathways. Wright argued that

variations in metabolic activity brought about by the heterozygous condition are likely to have little effect on the phenotype because enzymes are linked together in pathways so that the substrate of one enzyme is the product of another. Recessive mutations, when homozygous, may halt the activity of one enzyme and thus bring the entire pathway to a halt, producing a mutant phenotype. Heterozygotes, on the other hand, are likely to have only a reduction in activity of one enzyme which will be averaged out over the entire metabolic pathway, producing little phenotypic effect.

Molecular studies of dominance have extended Wright's ideas by exploring the kinetic structure of metabolic pathways and enzyme systems. The results obtained thus far tend to support the thrust of his hypothesis, and have established that the dominant phenotype seen in heterozygotes for a recessive allele can be explained without the need to invoke the existence of modifiers. Other work on organisms ranging from *Drosophila* to humans has determined the relationship between a mutant recessive allele, detectable changes in enzyme activity, and the phenotype. Reduction in enzyme activity is common to most recessive mutations in the heterozygous condition, with only negligible effects on the phenotype. Even reductions to 50% of normal enzyme activity in the heterozygote are not usually detectable at the phenotypic level. The example of phenylketonuria heterozygotes with 50% reduction in enzyme activity but normal phenotype is typical for human metabolic mutations. All such mutations are classified as recessive, leading to the speculation that the widespread occurrence of recessive mutations is a consequence of the kinetic organization and structure of enzyme networks and enzyme-directed pathways. *See* GENETICS; MOLECULAR BIOLOGY; MUTATION. Michael R. Cummings

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Donkeys and allies

Donkeys, asses, and mules are included in the family Equidae along with the horses and zebras. These are odd-toed ungulates and therefore belong to the mammalian order Perissodactyla.

Donkey. These animals, *Equus asinus* (see **illus.**), originated in Africa, and their only relatives in the wild state are found in Ethiopia and northwestern Somalia (Somaliland). They have been domesticated

The donkey, *Equus asinus*, sometimes referred to as a burro.

and are used principally as pack animals throughout Asia and Europe. The donkey was brought to the United States in 1868. They are hardy animals but better suited to hot climates since they are sensitive to cold. No race can be defined since they are the result of many generations of crossbreeding, but the European donkeys have certain markings which resemble the wild ass.

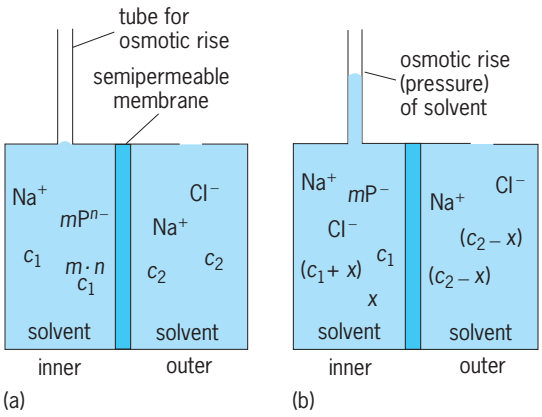
Ass. The wild ass of Asia (*E. hemionus*) is known by a variety of common names, and because of its extensive distribution it has been divided into a number of races or subspecies. In the central regions of Mongolia is the kulan (*E. hemionus*), the onager (*E. h. onager*) occurs in Persia and India, while the kiang (*E. kiang*), largest of all the races, is restricted to Tibet and the Sikkim. The Syrian wild ass (*E. h. hemippus*) is extremely rare and is believed to be extinct by many authorities. All of these animals are about 4 ft (1.2 m) high at the withers. They are the most abundant of the wild Equidae after the zebras.

The African wild ass (*E. asinus*) differs from the wild horse in having longer ears, a short tail which is only a tuft of hair, narrower hooves, and a shorter, more upright mane. This species is also becoming rare in its natural habitat, being found only between the Sudan and Somaliland.

Mule. This animal is a hybrid resulting from the cross of the male ass (*E. asinus*) and the mare or female horse (*E. caballus*). The mule is a surefooted animal, noted for its endurance, unexcitability, and ability to withstand hot weather. When the female ass is bred to a stallion, the cross is called a hinny. These crosses do not occur naturally and both mule and hinny are usually sterile. *See* MULE PRODUCTION; PERISSODACTYLA; ZEBRA. Charles B. Curtin

Donnan equilibrium

The distribution of diffusible ions on either side of a semipermeable membrane in the presence of macroions that are too large to pass through the membrane. The Donnan equilibrium thus is the result of (1) external constraints (boundary conditions) that



Schematic showing two compartments separated by a semipermeable membrane (*a*) before equilibrium and (*b*) at equilibrium. Originally, the left compartment contains *m* macroelectrolyte ions (P^{-n}) and its counterions (Na^+), in aqueous solution. The right compartment contains only NaCl in water. The membrane is permeable for Na^+ and Cl^- ions (and water) only.

enforce an unequal distribution of mobile ions and (2) a corresponding electrical potential on the membrane as a balance. This equilibrium is named for F. G. Donnan, the first to describe it and formulate the theory.

Typically, within a solution system that consists of two communicating compartments, the polyions of a macromolecular polyelectrolyte (a colloidal electrolyte) are confined to one compartment (for example, the inner one) by the semipermeable membrane, which allows the exchange of solvent and low-molecular-weight electrolytes by diffusion between the inner and outer compartment. Since electrical neutrality must obtain in either compartment, and since the macro-ions are confined to one side, the concentration of the diffusible ions in the two compartments cannot be the same (see **illus.**). The concentrations of the three ionic components (Na^+, Cl^-, P^{-n}) at the beginning of the experiment (illus. *a*) are: m, for P^{-n}, and $m \times n = c_1$ for Na^+ on the inside, and $Na^+ = Cl^- = c_2$ on the outside. After osmotic equilibrium has been reached and the membrane potential has been set up (illus. *b*), the concentrations are $(c_1 + x)$ for Na^+ on the inside, m unchanged for P^{-n}, and x for the Cl^-. On the outside, $Na^+ = Cl^- = (c_2 - x)$. The equilibrium condition is $(c_1 + x)x = (c_2 - x)^2$. At thermodynamic equilibrium, assuming for simplicity an electrolyte of which both ions are small, carry a charge of 1, and diffuse readily throughout the system in which they are dissolved, the chemical potential (activities) for the outer ionic pair must equal that of the diffusible ion pair on the inside. This condition is fulfilled when the product of the activities are equal, as shown in Eq. (1),

$$m_o^+ m_o^- \gamma_o^{\pm} = m_i^+ m_i^- \gamma_i^{\pm} \qquad (1)$$

where m_o and m_i are the molalities and γ_o and γ_i are the activity coefficients within the outer and inner compartments, respectively. In dilute solution, the molalities are approximately equal to the molarities, and the activity coefficients are practically equal to 1. The equilibrium of the solvent (whether the nondiffusible species is ionized or not) follows upon flux of the solvent into the (closed) cell that holds the macromolecular solute, until osmotic pressure equalizes the activities of the solvent in both cells. *See* ACTIVITY (THERMODYNAMICS).

Donnan membrane potential. The inequality of the individual ionic concentrations on either side of the membrane means that, at equilibrium, the membrane must carry a charge whose free energy is equal in magnitude and opposite in sign to the free energy that results from the unequal individual activities of the diffusible ions. This is shown in Eq. (2) for a negatively charged nondiffusible polyelectrolyte, where E is the potential difference. In other words, the presence of the nondiffusing species on one side of the membrane and the resulting differences in salt concentration enforce a polarization of the membrane (which is independent of the nature of the diffusible ions). Because of the equilibrium, the direct measurement of the electromotive force (membrane potential) by a pair of ion-reversible electrodes in each compartment would yield zero. However, the membrane potential can be measured by inserting two calomel reference electrodes into the inner and outer compartments while short-circuiting the two reversible electrodes, effectively bypassing the membrane. *See* ELECTROMOTIVE FORCE (EMF); FREE ENERGY; REFERENCE ELECTRODE.

$$E = \frac{RT}{F} \ln \frac{[Na_i^+]}{[Na_o^-]} = \frac{RT}{F} \ln \frac{[Cl_o^-]}{[Cl_i^-]} \qquad (2)$$

Observations. Equation (1) permits the determination of the concentration (activity) of the ions of the same charge as the macro-ions within the inside cell, which is equal to the activity of the ions that have diffused in from the outer cell (see illus.). It can also be seen that the osmotic pressure of a closed inner cell must be larger than the osmotic pressure that would be caused by the macro-ions alone, by an amount that corresponds to the number of ions that compensate the charges on the macro-ions, plus that of the ions that diffused in; that is, the osmotic pressure depends on the average molecular weight of the nonsolvent components of the inner cell. Therefore, if the salt concentration of the outer cell, and along with it the concentration in the inner cell, is raised, the contribution to the osmotic pressure by the macro-ions becomes relatively smaller and eventually becomes constant, amounting to the osmotic pressure of the macro-ions in their uncharged state. At the same time, the concentrations of the diffusible ions inside and outside approach one another. The same situation obtains in the absence of a membrane, when microscopic gel particles possess covalently bound ionized groups in their interior. Suspended microscopic gel particles, or dense polyelectrolyte coils, may thus carry a charge due to the effect of the Donnan equilibrium. *See* OSMOSIS.

The Donnan equilibrium is a frequent contributor to membrane potentials, but differential adsorption or, as in living systems, differences in the rates of passive or active ion transport through biological membranes are usually the main sources of membrane potentials of living cells.

If the concentration of the diffusible ions on the outside is kept near zero, as in the case of dialysis, the ions that were the counterions of the macroelectrolyte eventually diffuse out and become replaced by hydrogen ions (H^+), or hydroxide ions (OH^-); the original macro-micro-ion-salt (a salt containing ions of different sizes) has been hydrolyzed by the removal of the diffusible ions. *See* COLLOID; DIALYSIS; HYDROLYSIS; ION-SELECTIVE MEMBRANES AND ELECTRODES. F. R. Eirich

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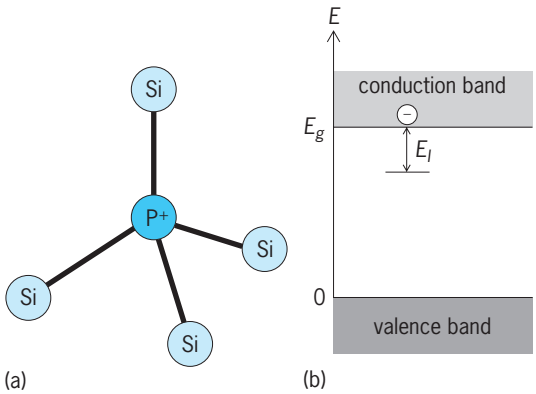
Donor atom

An impurity atom in a semiconductor which can contribute or donate one or more conduction electrons to the crystal by becoming ionized and positively charged. For example, an atom of column 15 of the periodic table, such as phosphorus, arsenic, or antimony, substituting for a regular atom of a germanium or silicon crystal is a donor because it has one or more valence electrons which can be detached and added to the conduction band of the crystal (see **illus.**). Donor atoms thus tend to increase the number of conduction electrons in the semiconductor. The ionization energy of a donor atom is the energy required to dissociate the electron from the atom and put it in the conduction band of the crystal. *See* ACCEPTOR ATOM; SEMICONDUCTOR. H. Y. Fan



Pentavalent donor atom, phosphorus (P), in the elemental semiconductor silicon (Si). (*a*) Phosphorus atom in a substitutional position, that is, replacing silicon, a tetravalent host atom, by completing the four tetrahedral covalent bonds with its nearest neighbor silicon atoms. This results in an extra electron not needed in this bonding scheme, making phosphorus positively charged. (*b*) Energy diagram showing the conduction band lying at an energy E_g above the valence bond. The fifth electron from subtitutional phosphorus is donated to the empty conduction band. The donor electron is bound to phosphorus via Coulomb attraction with an ionization energy E_I.

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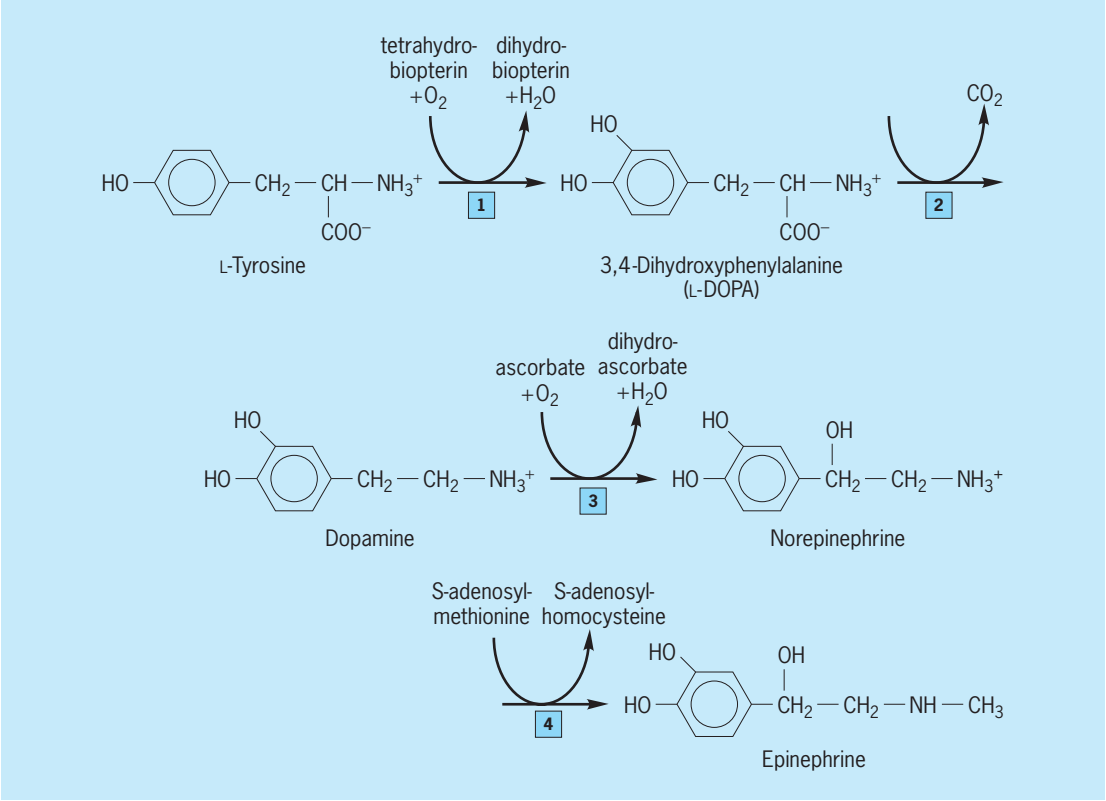
Dopamine

A catecholamine neurotransmitter that is synthesized by certain neurons in the brain and interacts with specific receptor sites on target neurons.

Biosynthesis. Catecholamine neurotransmitters—such as dopamine, norepinephrine, and epinephrine—are synthesized inside certain neurons from the common amino acid L-tyrosine. The biosynthetic pathway for catecholamine neuroransmitters has four steps (see **illustration**). (1) The enzyme tyrosine hydroxylase converts L-tyrosine to 3,4-dihydroxyphenylalanine (L-DOPA); this reaction also converts tetrahydrobiopterin and molecular oxygen to dihydrobiopterin and water. Tyrosine hydroxylase is found in all cells containing catecholamine neurotransmitters and is the regulatory, or rate-limiting, step in catecholamine biosynthesis. (2) The enzyme aromatic amimo acid decarboxylase converts L-DOPA to dopamine and carbon dioxide. In dopamine-producing cells, the dopamine is then stored in vesicles within the nerve terminal, which can fuse with the cell membrane to release dopamine into the synapse. (3) In neurons that produce norepinephrine and epinephrine, the enzyme dopamine beta-hydroxylase converts dopamine to norepinephrine; this reaction also converts ascorbate and molecular oxygen to dehydroascorbate and water. (4) In neurons that produce epinephrine (chromaffin cells), the enzyme phenylethanolamino *N*-methyltransferase converts norepinephrine to epinephrine; this reaction also converts *S*-adenosylmethionine to *S*-adenosylhomocysteine. *See* EPINEPHRINE.

Release from neuron. The release of neurotransmitter is controlled by a variety of factors, including the firing rate of the dopamine nerve cell (termed impulse-dependent release) and the release- and synthesis-modulating presynaptic dopamine receptors located on the dopamine nerve terminals. Since presynaptic dopamine receptors are sensitive to the cell's own neurotransmitter, they are called dopamine autoreceptors. Once released, dopamine also acts at postsynaptic receptors to influence behavior. The actions of dopamine in the synapse are terminated primarily by the reuptake of neurotransmitter into the presynaptic terminal by means of an active dopamine transporter. Dopamine may then be either repackaged into synaptic vesicles for rerelease or degraded by the enzyme monoamine oxidase. In addition, there is some extracellular inactivation. The dopamine transporter is an important



Biosynthetic pathway for catecholamine neurotransmitters.

site of action of the drugs cocaine and amphetamine. *See* AMPHETAMINE; COCAINE; MONOAMINE OXIDASE.

Dopamine neuronal systems. Although it was first thought that dopamine occurred only as an intermediate product formed in the biosynthesis of two other catecholamine neurotransmitters, norepinephrine and epinephrine, dopamine is now recognized as a neurotransmitter in its own right. Several distinct dopamine neuronal systems have been identified in the brain. These include systems within the hypothalamus and the pituitary gland; systems within the midbrain that project to a variety of cortical and limbic regions and basal ganglia; the retinal system; and the olfactory system. *See* BRAIN; NORADRENERGIC SYSTEM.

Role in behavior. The midbrain dopamine neurons which project to a variety of forebrain structures are critically involved in normal behavioral attention and arousal; abnormalities in the normal functioning of these systems have been implicated in a variety of disorders. For example, Parkinson's disease involves a degeneration of the midbrain dopamine neurons. This condition is often successfully treated by providing affected individuals with L-dopa, which is readily converted to dopamine in the brain. Attention deficit disorder, which is usually first diagnosed in childhood, is thought to involve dopamine systems, because the treatment of choice, methylphenidate, binds to the dopamine transporter and alters dopamine levels in the synapse. *See* PARKINSON'S DISEASE.

Drugs used to treat the major symptoms of schizophrenia are potent dopamine receptor antagonists. It is possible that certain types of schizophrenia are the result of increased activity in dopamine neuronal systems. In particular, it has been suggested that increased activity of either mesocortical or mesolimbic systems is likely to be involved. Although there is no doubt that neurobiological alterations underlie these disorders, a clear alteration in dopamine systems in the postmortem brain of schizophrenic individuals has not been conclusively demonstrated. A similar involvement of midbrain dopamine systems has been implicated in the multiple tic disorder Tourette's syndrome, which is treated, often successfully, with dopamine receptor antagonists. *See* NEUROBIOLOGY; SCHIZOPHRENIA.

Louis A. Chiodo; M. Todd Washington

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Doppler effect

The change in the frequency of a wave observed at a receiver whenever the source or the receiver is moving relative to each other or to the carrier of the wave (the medium). The effect was predicted in 1842 by C. Doppler, and first verified for sound waves

by C. H. D. Buys-Ballot in 1845 from experiments conducted on a moving train.

Acoustical Doppler effect. The Doppler effect for sound waves is now a commonplace experience: If one is passed by a fast car or a plane, the pitch of its noise is considerably higher in approaching than in parting. The same phenomenon is observed if the source is at rest and the receiver is passing it.

Quantitative relations are obtained from a galilean transformation between the system of the receiver, moving with velocity ($\vec{v}$) and that of the source, moving with velocity ($\vec{w}$) with respect to the medium, which may be air taken at rest. In the rest frame the wave propagates with the proper sound velocity given by Eq. (1), where λ and ν are the wavelength

$$|\vec{c}| = c = \lambda\nu \approx 333 \text{ m/s} = 1015 \text{ ft/s} \qquad (1)$$

and frequency measured in that system. Consider propagation of the wave in the direction of the vector ($\vec{c}$), forming angles α and β with the velocities ($\vec{v}$) and ($\vec{w}$) (**Fig. 1**). The frequency ν_R received is by definition the number of wavefronts (wave crests) passing the receiver per unit of time. This is given by calculating the difference between the sound velocity c and the component of the receiver velocity parallel to it $v_{\uparrow\uparrow}$ divided by the wavelength as in Eq. (2).

$$\nu_R = \frac{1}{\lambda}(c - v_{\uparrow\uparrow}) = \nu\left(1 - \frac{v_{\uparrow\uparrow}}{c}\right) = \nu\left(1 - \frac{v\cos\alpha}{c}\right) \qquad (2)$$

A moving source acts differently since it effectively changes the wavelength in the medium, as seen from the following argument: Let the source emit a wave crest at time $t = 0$, traveling during one period of the source oscillation, $T_s = 1/\nu_s$, a distance $c \cdot T_s$. The source follows the wave crest with its velocity component $w_{\uparrow\uparrow}$, parallel to ($\vec{c}$). Therefore it emits the next wave crest at a distance from the first wave crest given by Eq. (3), which is by definition the wave-

$$s = (c - w_{\uparrow\uparrow})T_s = \frac{c - w_{\uparrow\uparrow}}{\nu_s} = \lambda = \frac{c}{\nu} \qquad (3)$$

length λ in the medium. Hence (4) is obtained for the

$$\nu = \frac{\nu_s}{1 - (w_{\uparrow\uparrow}/c)} = \frac{\nu_s}{1 - [(w\cos\beta)/c]} \qquad (4)$$

frequency ν observed in the frame of the medium.

In both cases the transverse velocity components $v_{\uparrow\rightarrow}$ and $w_{\uparrow\rightarrow}$ have no effect since they run parallel to the wavefront. In the supersonic regime where w is greater than c, Eq. (4) holds only inside the Mach cone. *See* SUPERSONIC FLIGHT.

Combining Eqs. (2) and (4) relates the received frequency to the emitted frequency by the general relation of Eq. (5), which may be expanded up to second

$$\nu_R = \nu_s \frac{1 - (v_{\uparrow\uparrow}/c)}{1 - (w_{\uparrow\uparrow}/c)} \qquad (5)$$

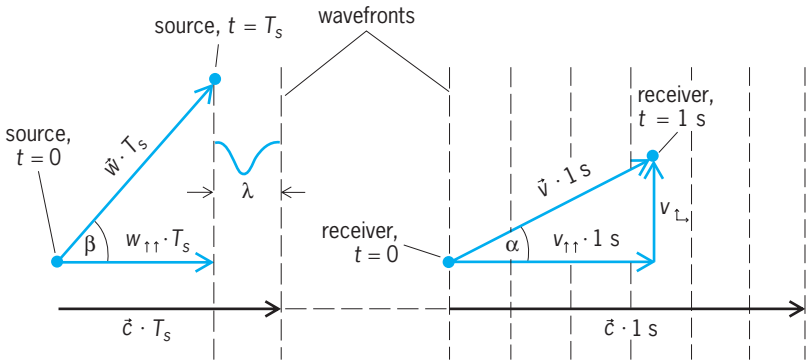

Fig. 1. Motions of source, receiver, and propagating sound that produce the acoustical Doppler effect.

order in the velocities, as in Eq. (6). The first-order

$$\nu_R = \nu_s\left(1 + \frac{w_{\uparrow\uparrow} - v_{\uparrow\uparrow}}{c} + \frac{(w_{\uparrow\uparrow} - v_{\uparrow\uparrow})w_{\uparrow\uparrow}}{c^2} + \cdots\right) \qquad (6)$$

term, the so-called linear Doppler effect, depends only on the relative motion between source and receiver, that is, on Eq. (7), where ($\vec{v}$) is the relative

$$u_{\uparrow\uparrow} = w_{\uparrow\uparrow} - v_{\uparrow\uparrow} \qquad (7)$$

velocity of the source with respect to the receiver, whereas the higher-order terms depend explicitly on the motion of the source relative to the medium. *See* SOUND.

Optical Doppler effect. The linear optical Doppler effect was first observed by J. Stark in 1905 from a shift of spectral lines emitted by a beam of fast ions (canal rays) emerging from a hole in the cathode of a gas discharge tube run at high voltage. Still, their velocity was several orders of magnitude below that of light in vacuum, c_0. This precluded a test of the second-order effect which would be sensitive to any kind of medium carrying light waves, the so-called light ether, if it existed. However, the precise interferometric experiments of A. A. Michelson and E. W. Morley (1887) which are closely related to the above considerations were sensitive to second-order effects. They showed clearly that the velocity of light is not bound to any ether, but is measured to be the same in any moving system. This result was a crucial check for A. Einstein's theory of special relativity (1905), which also makes a clear prediction for the optical Doppler effect with the help of a Lorentz transformation between source and receiver instead of a galilean transformation. This prediction is given by Eq. (8), where ($\vec{u}$) is the relative velocity, and ϑ

$$\nu_R = \nu_s \frac{\sqrt{1 - u^2/c_0^2}}{1 - [(u\cos\vartheta)/c_0]} \qquad (8)$$

is the angle between ($\vec{v}$) and the light propagation velocity ($\vec{c}$), both measured in the system of the receiver (**Fig. 2**). Equation (8) agrees to first order with Eq. (6), but differs in higher orders since it does not

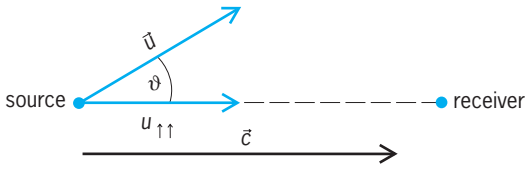

Fig. 2. Motions of source and propagating light, in the system of the receiver, that produce the optical Doppler effect.

contain the source velocity with respect to any ether, but depends entirely on the relative velocity ($\vec{v}$). *See* INTERFEROMETRY.

In contrast to the classical Doppler formula of Eq. (5), the exact relativistic optical Doppler effect of Eq. (8) also exhibits a shift for transverse motion ($\vartheta = 90^\circ$), namely by $\sqrt{1 - u^2/c_0^2}$. This transverse Doppler effect results in a lowering of the received frequency (a redshift), which is equivalent to the relativistic time dilatation in the moving system (a clock runs slower). A verification of the higher-order Doppler effect (and hence of special relativity) to a precision of 4×10^{-5} was obtained recently by two-photon spectroscopy (see below) on a beam of fast neon atoms. Since the first-order Doppler effect cancels in this method, the experiment yielded directly the redshift of the atomic transition in question. *See* RELATIVITY.

Applications. The Doppler effect has important applications in remote-sensing, high-energy physics, astrophysics, and spectroscopy.

Remote sensing. Let a wave from a sound source or radar source, or from a laser, be reflected from a moving object back to the source, which may itself move as well. Then a frequency shift $\Delta\nu$ is observed by a receiver connected to the source, which is given to first order by Eq. (9), where $u_{\uparrow\uparrow}$ is given by Eq. (7) and

$$\Delta\nu = \frac{2\nu_s u_{\uparrow\uparrow}}{c_0} \tag{9}$$

c_0 is the propagation speed of sound or electromagnetic radiation. The factor 2 is due to reflection. The measurement of $\Delta\nu$ is easily carried out with high precision, and provides an excellent means for the remote sensing of velocities of any kind of object, including cars, ships, planes, satellites, flows of fluids, or winds. *See* DOPPLER RADAR; REMOTE SENSING.

High-energy physics. Consider a head-on collision of a laser beam, with wavelength λ of approximately 6×10^{-7} m, with a beam of electrons of very high energy, say 5×10^{10} eV, which is therefore extremely relativistic with v/c approximately equal to $1 - 0.5 \times 10^{-10}$. Then the same considerations as in the case of remote sensing, but calculated relativistically, lead to the conclusion that the light, backscattered by the Compton effect, is changed from visible photons into high-energy gamma rays with a wavelength of approximately 4×10^{-17} m and a quantum energy $h\nu$ (where h is Planck's constant) of approximately 3.2×10^{10} eV. The electron has thus transferred most of its kinetic energy to the light quantum. Such experiments have been performed at high-energy accelerators. *See* COMPTON EFFECT.

Astrophysics. The light from distant stars and galaxies shows a strong Doppler shift to the red, indicating that the universe is rapidly expanding. However, this effect can be mixed up with the gravitational redshift that results from the energy loss which a light quantum suffers when it emerges from a strong gravitational field. *See* COSMOLOGY; GRAVITATIONAL REDSHIFT; REDSHIFT.

The Doppler width and Doppler shift of spectral lines in sunlight (Fraunhofer lines) are important diagnostic tools for the dynamics of the Sun's atmosphere, indicating its temperature and turbulence. *See* ASTRONOMICAL SPECTROSCOPY; SUN.

Doppler-free spectroscopy. Spectral lines from atomic or molecular gases show a Doppler broadening, due to the statistical Maxwell distribution of their velocities. Hence the spectral profile, that is, the intensity I as a function of frequency ν, is a gaussian centered around ν_0, given by Eq. (10), with full halfwidth given by Eq. (11), where k is the Boltzmann constant, T

$$I(\nu) = I(\nu_0) \exp - \left(\frac{\nu - \nu_0}{\delta\nu}\right)^2 \tag{10}$$

$$\delta_{1/2}(\nu) = (2 \ln 2)\delta\nu = (2 \ln 2)\frac{\nu_0}{c_0}\sqrt{\frac{2kT}{m}} \tag{11}$$

is the thermodynamic temperature in kelvins, and m is the mass of a molecule. The Doppler width in Eq. (11) is of order $10^{-6}\ \nu_0$ and exceeds the natural linewidth, given by Eq. (12), by orders of magnitude.

$$\delta\nu_{\text{nat}} = \frac{1}{2\pi\tau} \tag{12}$$

Equation (12) is the ultimate limit of resolution, set by the decay time τ of the quantum states involved in the emission of the line. Thus, precision spectroscopy is seriously hindered by the Doppler width. *See* LINEWIDTH; SPECTROSCOPY.

Around 1970, tunable, monochromatic, and powerful lasers came into use in spectroscopy and resulted in several different methods for completely overcoming the Doppler-width problem, collectively known as Doppler-free spectroscopy. Lasers also improved the spectroscopic sensitivity by many orders of magnitude, ultimately enabling the spectroscopy of single atoms, and both improvements revolutionized spectroscopy and its application in all fields of science. The most common Doppler-free methods are discussed below.

1. *Atomic beams.* An atomic beam, emerging from an oven, well collimated and intersecting a laser beam at right angles, absorbs resonant laser light with the Doppler width suppressed by a factor of $\sin\vartheta$, where ϑ is the collimation angle (**Fig. 3**). This straightforward method was already used before the advent of lasers with conventional light sources. *See* MOLECULAR BEAMS.

2. *Collinear laser spectroscopy on fast beams.* Another way to reduce the velocity spread in one direction is by acceleration. Consider an ion emerging from a source with a thermal energy spread δE, of the order of the quantity kT, or approximately 0.1 eV,

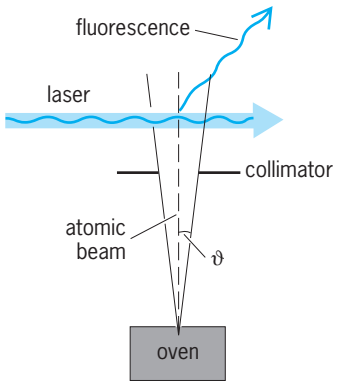

Fig. 3. Doppler-free atomic-beam spectroscopy on thermal atomic beams.

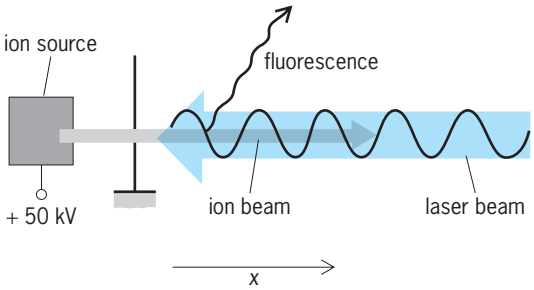

Fig. 4. Collinear laser spectroscopy on accelerated beams.

and accelerated by a potential of, say, 50 kV along the x direction (**Fig. 4**). The kinetic-energy spread in this direction remains constant, however, and may be expanded as in Eq. (13), where the overbars represent

$$\delta E \approx kT \approx \delta\left(\frac{m}{2}v_x^2\right) \approx m\overline{v}_x \cdot \delta v_x$$
$$\approx \frac{mc_0^2}{\nu_0^2} \cdot \overline{\Delta\nu} \cdot \delta\nu = \text{constant} \qquad (13)$$

average values. Since the average velocity $(\overline{u_x})_x$ has been increased by a factor of 700 in the example given, its spread δv_x, and hence the Doppler width with respect to a collinear laser beam, is reduced by the same factor. Consequently, the product of the Doppler shift $(\overline{\Delta\nu})$ of the fast beam and its Doppler width $\delta\nu$ is a constant of the motion. This method is very versatile and extremely sensitive.

3. *Laser cooling.* The Doppler effect has been used to cool down the thermal motion of atoms or ions by absorption of laser light. For instance, an ion may be trapped in a vacuum by electromagnetic fields (**Fig. 5**). The radial motion is stabilized by a magnetic field B, causing small cyclotron orbits of the ion, whereas axial motion is stabilized by applying a repelling electric potential to the electrodes on the axis. A laser excites the ion at the low-frequency side of the Doppler profile of a resonance line, thus selecting excitations at which the ion counterpropagates to the light. Therefore the ion is slowed down, since it absorbs the photon momentum $h\nu/c$ (where h is Planck's constant) which is opposite to its own. After the ion emits fluorescent radiation, the process is repeated until the ion comes almost to rest.

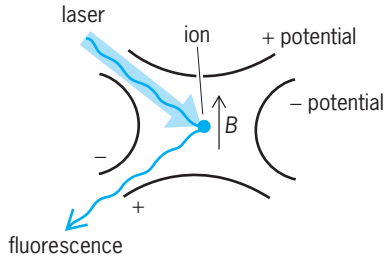

Fig. 5. Laser cooling of a trapped ion.

The repelling force of the cooling process has also been used to confine a cloud of free atoms in vacuum at the intersection volume of three mutually orthogonal pairs of counterpropagating lasers. Thermodynamic temperatures of the order of 10^{-4} kelvin have been reached in such systems. *See* LASER COOLING.

4. *Saturation spectroscopy.* Whereas the velocity distribution of the atoms is manipulated in the above examples, in the following ones this distribution is left untouched, but Doppler broadening is suppressed through the nonlinear optical interaction of the photons with the atoms. The field of nonlinear optics can be roughly characterized by the condition that the atomic system interacts with more than one light quantum during its characteristic decay time τ. In saturation spectroscopy, two almost counterpropagating laser beams of the same frequency ν are used, a strong pumping beam and a weak one (**Fig. 6**). The

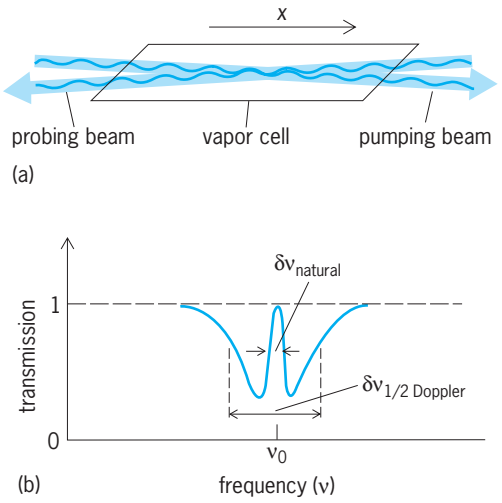

Fig. 6. Saturation spectroscopy. (*a*) Experimental configuration. (*b*) Transmission of probing beam.

pumping beam is strong enough to saturate the excitation of those atoms in the vapor which match the resonance condition, that is, which compensate the frequency offset from the center $\nu - \nu_0$ by the proper velocity component, given by Eq. (14). Satu-

$$v_x = c(\nu - \nu_0) \qquad (14)$$

ration means that the excitation is fast as compared to the decay time, leading to an equal population of

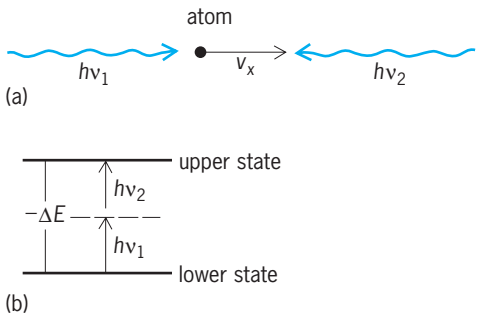

Fig. 7. Two-photon spectroscopy. (*a*) Relation of atom's motion to photons from counterpropagating laser beams. (*b*) Energy-level diagram.

the lower and the upper atomic states. Such a system is completely transparent. The probing beam feeds on atoms with exactly the opposite velocity component given by Eq. (15). It is too weak to sat-

$$-v_x = c\,(\nu - \nu_0) \tag{15}$$

urate, and is absorbed in the vapor except at the center frequency $\nu = \nu_0$, where it matches with the saturated part of the velocity spectrum. Resolutions of $\delta\nu/\nu_0 < 10^{-10}$ obtained in this way make possible improved values of fundamental constants and of the standards of length and time. *See* FUNDAMENTAL CONSTANTS; LENGTH; LIGHT; PHYSICAL MEASUREMENT; RYDBERG CONSTANT.

5. *Two-photon spectroscopy.* In this method, two counterpropagating laser beams are operated at half the transition frequency ν_0 of the atomic system. Then Bohr's quantum condition, that the transition energy ΔE equals $h\nu_0$, can be fulfilled by simultaneous absorption of two photons, say, one from each beam (**Fig. 7**). Including the first-order Doppler effect, the energy balance is then given by Eq. (16).

$$h\nu_1 + h\nu_2 = h\nu\left(1 + \frac{v_x}{c}\right) + h\nu\left(1 - \frac{v_x}{c}\right)$$
$$= 2h\nu = h\nu_0 \tag{16}$$

Thus, the first-order Doppler effect cancels completely, and the two-photon resonance occurs sharply at half the transition frequency. *See* ATOMIC STRUCTURE AND SPECTRA; LASER SPECTROSCOPY; NONLINEAR OPTICS.

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Doppler radar

A radar system used to measure the relative velocity of the system and the radar target. The operation of these systems is based on the fact that the Doppler frequency shift in the target echo is proportional to the radial component of target velocity. *See* DOPPLER EFFECT.

Airborne vehicular systems. Airborne systems are used to determine the velocity of the vehicle relative to the Earth for such purposes as navigation, bombing, and aerial mapping, or relative to another vehicle for fire control or other purposes. Ground or ship equipment is used to determine the velocity of vehicular targets for fire control, remote guidance, intercept control, traffic control, and other uses.

The Doppler frequency shift Δf is an extremely small fraction of the transmitter frequency f. It is given by the equation shown below,where V is the

$$\Delta f = \frac{2Vf}{C}\cos\gamma$$

relative speed, C is the speed of signal propagation, and γ is the angle between the velocity and the direction of propagation (**Fig. 1**). The only practical way to measure the frequency shift is by adding the echo signal to a reference signal derived from the transmitter and observing the difference, or beat frequency. Some means of obtaining coherent detection is required.

Practical techniques have been devised for obtaining the requisite coherence in continuous-wave (CW), pulsed, and frequency-modulated transmission systems. For a discussion of techniques similar to those in Doppler radar *see* CONTINUOUS-WAVE RADAR; MOVING-TARGET INDICATION.

Doppler navigation radar is a type of airborne Doppler radar system for determining aircraft velocity relative to the Earth's surface. Such a Doppler velocity sensor (**Fig. 2**) consists of at least the following elements: transmitter, antenna, receiver, Doppler frequency measuring device, and output signal generators or displays. It is generally used with a navigation computer. *See* AIRBORNE RADAR.

The signal from a single beam can provide only the velocity component in the direction of that beam. Complete velocity determination requires, therefore, the use of at least three beams. Most systems use four beams for symmetry (**Fig. 3**).

To relate the beam directions, and hence the

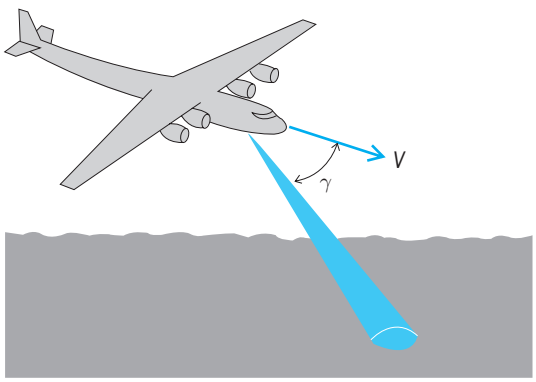

Fig. 1. Doppler frequency measurement geometry.

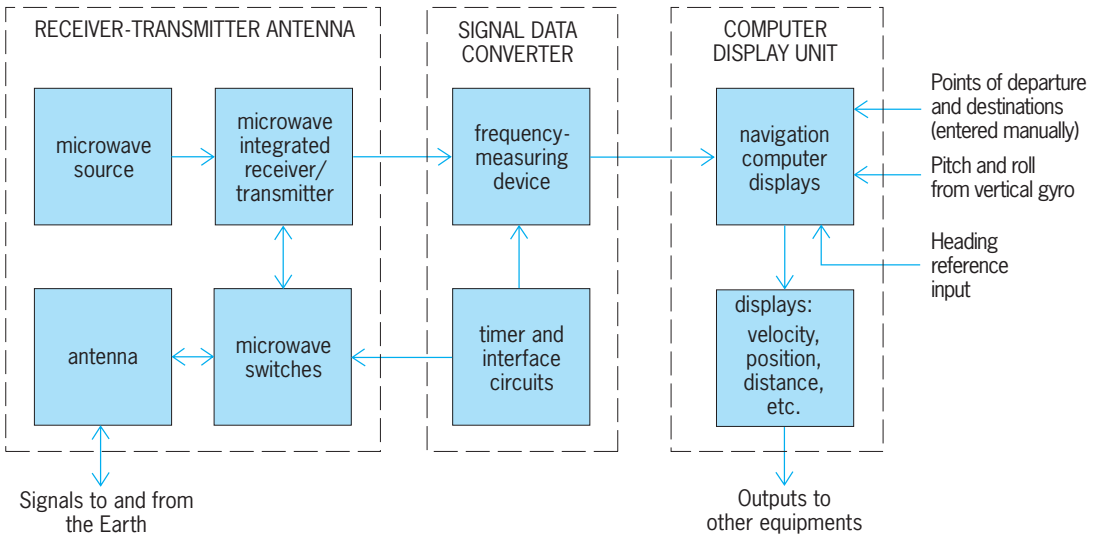

Fig. 2. Block diagram of aircraft Doppler navigation system, the AN/ASN-128.

measured velocity to an Earth-oriented coordinate system, a vertical reference must be provided. In newer systems the antennas are fixed to the aircraft. The Doppler frequencies and the vertical data, such as roll and pitch, are fed to a computer. Its outputs are electrical signals or displays representing the components of velocity. The components commonly displayed are those along-heading, across-heading, and vertical.

Various types of antennas have found use in Doppler navigation radar. These include paraboloids, microwave lenses, and linear and planar arrays. Since both volume and radome cutout area should be small, various techniques are employed to enable each antenna to form (simultaneously or sequentially) more than one beam. Pencil beams of 3–5° width are used. The beams are directed 15–25° from the vertical. Larger values result in insufficient echo power over water. *See* ANTENNA (ELECTROMAGNETISM).

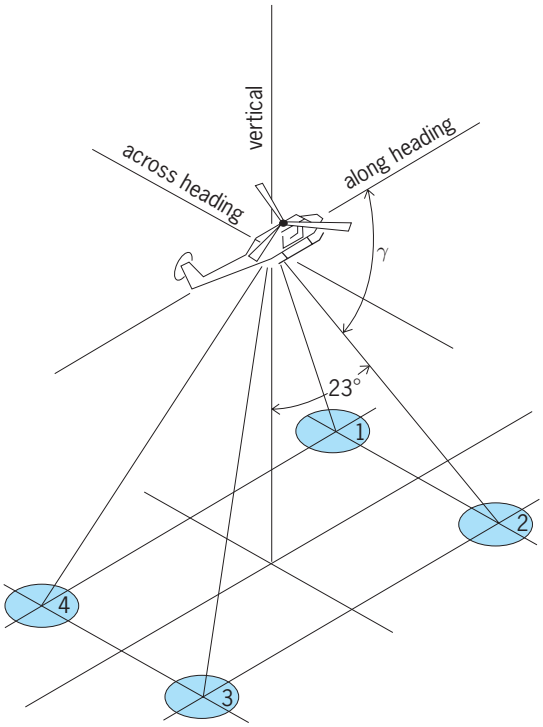

Fig. 3. Typical antenna beam arrangement.

Continuous-wave systems are coherent and are theoretically the most efficient. The chief difficulty is control of leakage of spurious signals from transmitter to receiver.

Pulsing enables the receiver to be rendered insensitive during transmission, thereby avoiding leakage signals. Coherence is achieved either by driving a transmitting power amplifier from a CW oscillator or by mixing at the detector a pair of pulse echoes received over different propagation paths.

A preferred technique is to employ sinusoidal frequency modulation. A sideband of the detected beat between echo and transmitter signal is used. Modulation index and rate and the sideband order are chosen such that echoes from nearby objects are rejected, while those from distant objects are accepted. Leakage noise is reduced at the expense of lowered efficiency.

An example of a system (**Fig. 4**) is the U.S. Army's standard airborne Doppler navigator which bears the nomenclature AN/ASN-128. It weighs 28 lb (12.7 kg), requires 89 W of power, and operates at 13.325-GHz frequency. In addition to velocities, outputs include present position, distance to go, desired track and bearing to destination, and steering signals. For helicopter use a steering hover indicator can also be provided. The system uses all-solid-state electronics and a printed grid planar array antenna. The velocity accuracy specifications are around a quarter of a percent of ground speed (and 0.2 knot or 0.1 m/s when hovering). The navigation accuracy specification is 1.3% of distance traveled when using a 1° accuracy heading reference.

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Fig. 4. AN/ASN-128 Doppler navigation system. (*a*) Receiver-transmitter antenna. (*b*) Computer display unit. (*c*) Steering hover indicator. (*d*) Signal data converter. (*Singer Co., Kearfott Division*)

Meteorological systems. Doppler radars measure the radial velocity of targets by processing the Doppler frequency shift associated with this velocity, and therefore they are well suited for the observation of the motion of raindrops in precipitation systems. However, precipitation has the form of a distributed target, that is, numerous independent scatterers distributed in space. The backscattered signal, selected by a sampling circuit called the range gate, is thus composed of a large number of separate scattering amplitudes, each having a Doppler frequency shift associated with the radial velocity of a particular scatterer. Therefore, a spectrum of frequency shifts (Doppler spectrum) is observed.

To obtain the Doppler spectrum, the radar signal sampled at each range gate must be submitted to a Fourier transform operation. This is typically done by digitizing that signal and implementing fast Fourier transform algorithms. Spectral moments such as mean Doppler velocity and spectrum variance are also calculated. By using minicomputers equipped with specialized cards or hardware systems based on integrated circuits, fast Fourier transforms can be performed continuously at a large number of range gates whose spacing essentially matches the radar range resolution (a few hundred meters). However, much simpler algorithms are widely used to obtain the spectral moments by direct processing of time signals. These algorithms, called pulse-pair algorithms, proceed by computing the signal autocovariance whose argument is the mean Doppler frequency normalized to the frequency sampling domain established by the radar pulse repetition frequency. *See* ALGORITHM; HARMONIC ANALYZER.

Design. By using suitable phase coherence and signal-processing techniques, the Doppler shift can be extracted from the radar signal and interpreted in terms of the target's radial velocity. The sampling of the backscattered signal at each range gate is done at the radar pulse repetition frequency, and therefore unambiguous Doppler frequency observations are made only in frequency range allowed by the sampling rate. Since the Doppler frequency shift for a given radial velocity is inversely proportional to the radar wavelength, the unambiguous velocity domain for a given pulse repetition frequency increases in proportion to this wavelength. For example, a 1-kHz pulse repetition frequency (corresponding to an unambiguous range of about 300 km or 186 mi) yields an unambiguous velocity domain of ± 25 m/s (56 mi/h) at a wavelength of 10 cm compared to ± 8 m/s (18 mi/h) at a 3.2-cm wavelength. For long-range precipitation monitoring, it is thus advantageous to use a long radar wavelength. However, the radar beamwidth for a given antenna size increases with radar wavelength, so that, for an antenna of practical size, the choice of wavelength is limited to 10 cm or less. A typical 10-cm-wavelength Doppler radar such as the one proposed in the NEXRAD (next generation radar) program is equipped with a 10-m (33-ft) dish antenna, producing a 0.7° beamwidth. For that radar, a 2-kHz pulse repetition frequency unambiguous range of 150 km or 93 mi) yields a ± 50 m/s (112 mi/h) unambiguous velocity domain. By using velocity dealiasing algorithms that resolve first- and second-order velocity ambiguities, the effective velocity domain can be increased by ± 100 m/s (224 mi/h), which practically covers all the possible wind values encountered in precipitation systems.

Single radar. Conventional meteorological radars are being gradually replaced by Doppler radars, such as those of the NEXRAD program implemented by the U.S. Weather Service. In the new radars, signal processing, data treatment, and data display are performed by a computer dedicated to the system. An example of the usefulness of a single Doppler radar is the measurement of mean wind field as a function of altitude, assuming uniform wind conditions in the area scanned by the radar beam, from processing of radial velocity-range profiles acquired by scanning the radar beam continuously in azimuth. The NEXRAD network will also aid in the detection and monitoring of severe storms and tornadoes and the tracking of hurricanes, and in specifying conditions for wind shear or microburst occurrence in convective storms in an effort to improve air traffic safety.

Multiple radars. The joint operation of two or more Doppler radars simultaneously observing the same convective storm enables the measurement of the three components of the velocity of precipitation particles. While the horizontal motion of the

precipitation particles is identified with horizontal air motion, particle vertical velocity is the sum of two terms: fall speed of the particle in still air (terminal velocity) and vertical air velocity. If an estimate of particle terminal velocity is made (from radar reflectivity, for instance), then the three components of air velocity can be derived. The assumption that air compressibility is negligible also offers a means of deriving vertical air velocity from the divergence of the horizontal velocity field. Multiple Doppler radar networks have been used extensively in convective storm research and have contributed to rapid advances in the understanding of convective storm development. *See* CALCULUS OF VECTORS; FLUID-FLOW PRINCIPLES.

Airborne radars. Meteorological Doppler radars have been installed aboard airplanes for the purpose of measuring wind fields in storm systems which cannot be reached by ground-based radars. The simplest method of operation relies on the radar beam being perpendicular to the airplane course and scanning slightly in the vertical plane. One of the most striking examples of this method is the probing of the wind field inside a hurricane. The hurricane evolves slowly and thus can be probed from different directions by the radar in a time interval during which its evolution is negligible. Several radial-velocity "snapshots" of the same motion field can then be combined to obtain the horizontal motion field of different altitude levels. This method has been used primarily as a research tool and has produced unique observations of the kinematics of hurricanes which will ultimately help to improve their modeling and trajectory prediction. *See* HURRICANE.

Millimeter-wave radars. Centimeter-wave radars are excellent tools for the probing of precipitation, but they fail to detect and observe clouds, in which particle size is below approximately 100 micrometers. Since the amount of energy returned (backscattering) by cloud droplets increases with the inverse fourth power of the radar wavelength, millimeter-wave radars can overcome this weakness. For example, the intensity of the echo returned by the same cloud droplet at a 3.2-mm wavelength is 10,000 times greater than that at a 3.2-cm wavelength. Modern development of millimeter-wave technology has produced Doppler radars operating at very short wavelengths (2–3 mm). These radars are excellent tools for the probing of cloud kinematics and liquid water, and have already produced results applicable to modeling of small cumulus clouds, for instance. They are also useful for continuous monitoring of cloud layers by providing observations of cloud water and vertical velocity. Since they produce very narrow beams with small antenna size and because the enhanced backscattering drastically reduces the need for transmitted power (and thereby power supply), millimeter-wave Doppler radars are excellent candidates for installation aboard satellites in an effort to observe motion in precipitation systems at a global scale. *See* CLOUD PHYSICS.

Doppler lidars. Doppler techniques have been applied to optical lidars (light detection and ranging). These systems operate in a manner similar to pulse Doppler radars by sending bursts of optical energy. The Doppler lidars now being used for meteorological research operate at a 10.6-micrometer wavelength. Because of the strong attenuation of infrared radiation by hydrometeors, their use is restricted to observing cloud boundaries or aerosol and pollution layers. Instead of relying on the pulse-to-pulse coherence type of operation common to pulse Doppler radars, the Doppler frequency is extracted in time intervals equal to the transmitted pulse duration and sampled from the received signal associated with one transmitted light pulse. This processing is repeated independently for several light pulses and associated echoes, with the average of these estimates calculated so as to improve the accuracy of the measurements. Only mean Doppler frequency can be extracted by using time-domain algorithms similar to those used in the pulse-pair technique discussed above, but designed for the computation of signal autocovariance within a time interval equal to a transmitted pulse duration.

Doppler lidars were initially implemented as ground-based systems operated in a vertically pointing beam mode, but they were later installed aboard airplanes. This method of operation has produced wind field measurements in the lower troposphere based on observing aerosol motion. Their use aboard satellites for measurement of the wind field has been proposed in regions where aerosol density is expected to produce sufficient backscattering intensity. *See* LIDAR; METEOROLOGICAL RADAR; RADAR; RADAR METEOROLOGY.

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Doppler VOR

A very high-frequency omnidirectional radio range (VOR) employing the Doppler principle. The bearing accuracy achieved by users when VOR beacons are installed in the vicinity of obstructions, or when aircraft using VOR signals fly over mountainous terrain, is deteriorated by reflections (site and enroute errors). The Doppler VOR solves this problem by two fundamental principles maintaining full compatibility of the radiated information with existing airborne receivers. It uses a wide-base antenna array (**Fig. 1**) for suppressing the effects of multipath propagation and the Doppler principle for determination of bearing. According to the Federal Radionavigation Plan of 2001, approximately half of the existing VOR stations in the United States will be phased out beginning in 2010. However, Doppler VOR will be installed

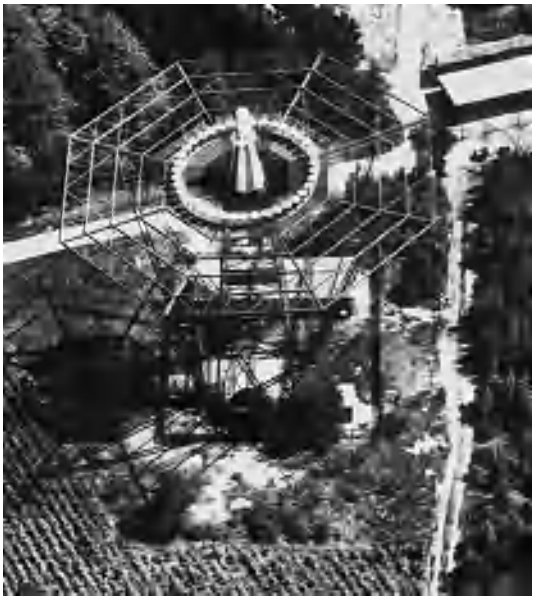

Fig. 1. Collocation of Doppler VOR and TACAN beacons installed on 100-foot high (30-m) tower with a counterpoise of 119-ft (36-m) diameter in forest near Frankfurt, Germany. Circular array consists of 39 loop antennas fed by solid-state commutator. Center carrier radiator is covered with plastic dome, which supports TACAN antenna.

in those remaining VOR sites where performance is degraded by multipath propagation. *See* VOR (VHF OMNIDIRECTIONAL RANGE).

Disturbances of VOR signals received by aircraft that are caused by multipath propagation can be reduced by wide-aperture systems because of integration or averaging between the information transferred over the direct and the reflected propagation path. In other words, when the reflections are weaker than the direct path signal, their effect tends to average out over intervals that are long compared to the multipath delay. Wide-base systems, however, have the drawback that their directional information becomes ambiguous as soon as the aperture is larger than half a wavelength, approximately.

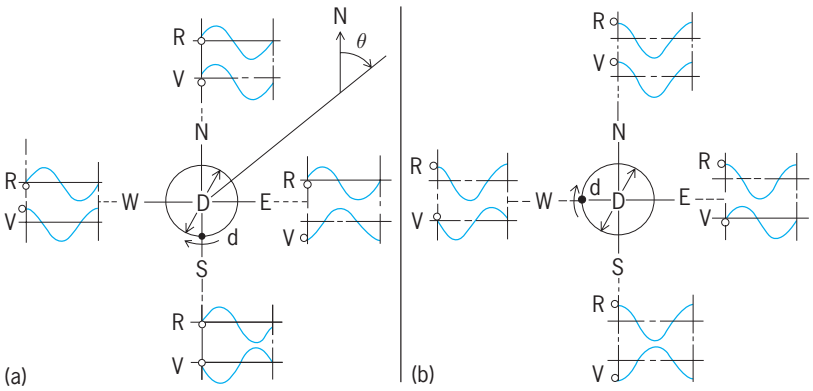

Fig. 2. Principle of Doppler VOR beacon. Reference signal R and bearing signal V which is generated by dipole d are shown as they would be measured in aircraft to the north, east, south, and west of the beacon. (*a*) Dipole d at 180°. (*b*) Dipole d at 270°. Phase difference between R and V gives bearing θ.

By applying the Doppler principle, one can overcome these limitations. A dipole rotating eccentrically on an orbit periodically changes the distance between transmitter and receiver. This periodic change of distance modulates the frequency of the radiated carrier signal by a deviation f_D according to the equation below. Here D is the diameter of the array and

$$f_D = \frac{\text{velocity}}{\text{wavelength}} = \frac{\pi D f_{\text{rot.}}}{\lambda}$$

$f_{\text{rot.}}$ is the number of rotations, or revolutions, per second. The resulting Doppler effect produces a sinusoidal frequency modulation (FM) of the carrier frequency; the phase of this FM contains the bearing information (variable phase).

In 1958 the U.S. Civil Aeronautics Agency constructed the first experimental Doppler VOR station in Indianapolis, Indiana. A nondirectional antenna radiates a radio-frequency (rf) carrier, amplitude-modulated with the rotational frequency, thus transmitting the reference information (**Fig. 2**). Another frequency, spaced by 9960 Hz from the carrier, corresponding to the subcarrier frequency of the conventional VOR, is radiated by an antenna which is in the form of a circle around the carrier antenna located in the center. The pattern generated by this antenna appears to rotate. Actually this circular motion is simulated by sequential feeding of radiators (Alford loops) arranged in a circular array. In contrast to the conventional VOR, the roles of the reference information (R) and the bearing information (variable signal V) are exchanged. This does not affect the existing airborne receiver functions.

For a bearing of 90° (aircraft east of ground station), for example, the phase difference is always 90°. When the reference signal for this bearing in Fig. 2*a* passes through zero, the variable signal is at its minimum because the dipole moves away from the observer in the east. When the dipole has moved by 90° on the circle, Fig. 2*b*, the variable signal reaches zero in the east because the motion is perpendicular to the east direction. Meanwhile the reference signal has increased to its maximum value. This mode of operation corresponds to single-sideband (SSB) modulation where the carrier is amplitude-modulated with the reference signal and its sideband is frequency-modulated with the variable signal which is determined by the instantaneous radiator position.

Certain disadvantages of this arrangement can be eliminated by the double-sideband (DSB) Doppler VOR. Here, the upper and lower sideband, spaced by 9960 Hz from the carrier, respectively, feed two radiators which are located diametrically opposite on the circular antenna array. Both sideband radiations are moved in steps from antenna to antenna along the circle. The signals radiated are now almost identical with those from conventional VOR ground stations. The carrier is radiated from the central antenna, as in the case of the SSB system.

The alternating-sideband (ASB) method is a variant

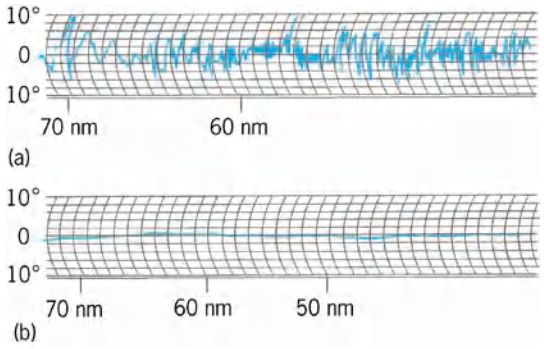

Fig. 3. Comparison of radial flight recordings. (*a*) From VOR installation. (*b*) From DVOR installation. Both experimental beacons were located at same unfavorable site. Plane is flying constant course at bearing of 133° from beacons at 4000-ft (1.2-km) altitude. Improvement of Doppler VOR is obvious. Radial flight recordings revealed for conventional VOR course some fluctuations which exceeded ICAO limits; maximum deviation of DVOR was approximately 0.5°.

of the double-sideband principle. This beacon is provided with an odd number of antennas on a circle; therefore no antenna is diametrically opposite to another. Upper and lower sideband interrupted carrier-wave trains are alternately fed to the nearly opposite antennas so that the energy travels at a constant speed around the circle. This system uses only half as many antennas as the array of a DSB system, and some difficulties are avoided concerning mutual coupling between adjacent antennas. In contrast to an SSB Doppler VOR which uses an antenna spacing of less than half a wavelength, the ASB array employs elements spaced by two-thirds a wavelength or more.

The Doppler VOR presents two essential advantages over conventional VOR: the improvement of absolute accuracy in good and poor terrain, and the spectacular radial stability and smoothness even over mountains (**Fig. 3**).

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Dormancy

In the broadest sense, the state in which a living plant organ (seed, bud, tuber, bulb) fails to exhibit growth, even when environmental conditions are considered favorable. In a stricter context, dormancy pertains to a condition where the inhibition of growth is internally controlled by factors restricting water and nutrient absorption, gas exchange, cell division, and other metabolic processes necessary for growth. By utilizing the latter definition, dormancy can be distinguished from other terms such as rest and quiescence which reflect states of inhibited development due to an unfavorable environment.

This article deals with dormancy in a wide array of plant structures. However, the main focus centers on the regulation of seed and bud dormancy. This does not preclude the fact that the plant kingdom consists of a vast collection of other specialized plant parts whose dormancy may be governed by more refined systems than those that fall within the scope of this article.

Physical influences. Physically induced dormancy can be separated into two distinct classes, based on external conditions imposed by the environment (light, temperature, photoperiod) and restraints induced by structural morphology (seed-coat composition and embryo development).

External conditions. The physical environment plays a key role in dormancy induction, maintenance, and release in several plant species.

1. *Temperature.* The onset of dormancy in many temperate-zone woody species coincides with decreasing temperature in the fall. However, it is the chilling temperature of the oncoming winter which is more crucial, particularly in regard to spring budbreak. The range of chilling requirements varies considerably, not only between different species, but also within a given species. In peach (*Prunus persica*), for example, chilling requirements among various cultivars range from 350 to 1200 h of exposure to temperatures below 45°F (7°C). Different portions of a plant can also differ in their particular chilling requirements, as in the case of fruit crops (for example, peach, apple, and cherry) where floral buds break much earlier than those which give rise to the new foliage.

Seeds of many diverse plant species, such as birch (*Betula* spp.), ash (*Fraxinus excelsior*), and grape (*Vitis vinifera*), also require a certain period of chilling, often coupled with adequate moisture to alleviate a dormant state. This process, which utilizes both cold temperature and moisture, is termed stratification and relates to the manipulation of endogenous factors affecting dormancy.

2. *Light duration and quality.* Possibly the single most important environmental variable affecting dormancy is day length or photoperiod. In plants capable of dormant bud production, such as temperate-zone woody trees, long photoperiods exceeding 12 h favor a continuation of growth, whereas short day lengths usually trigger dormancy. The same photoperiodic response occurs in tuberous crops, such as the Jerusalem artichoke (*Helianthus tuberosus*) and the potato (*Solanum tuberosum*), where short days induce tuber formation which eventually results in a dormant condition.

In other plants (for example, onion bulbs) long, hot days during summer droughts actually induce dormancy. Light quality and intensity also play an important role in dormancy in addition to day length. In certain lettuce cultivars there is a specific requirement for red light to break dormancy and induce germination. *See* PHOTOPERIODISM.

3. *Water and nutrient status.* Dormancy is affected by the availability of water and nutrients as demonstrated by many grasses, desert species, and subtropical fruits which go into dormancy when

confronted by drought or lack of soil fertility. Withholding of mineral fertilization is one practice utilized to induce tuberization and dormancy in tuberous crops. Conversely, continuous fertilization can postpone dormancy in several ornamental shrubs. In the case of dormant peach seeds, nitrogen and sulfur-containing compounds such as thiourea induce germination. *See* FERTILIZER; PLANT MINERAL NUTRITION; PLANT-WATER RELATIONS.

4. *Environmental interactions.* Several of the factors previously discussed do not simply act independently, but combine to influence dormancy. The temperature range for budbreak and seed germination, for instance, is regulated by day length. In certain situations long days actually replace the need for low temperatures in inducing seed germination. In other situations optimal temperatures or lighting regimes remove dormancy in seeds even under unfavorable moisture conditions.

Physical constraints. Examples of dormancy imposed by physical restrictions are most evident in the structural morphology of dormant seeds. These restrictions specifically pertain to the physical properties of the seed coat and developmental status of the embryo.

1. *Seed-coat factors.* The seed-coat material surrounding embryos of many plants consists of several layers of tissue, termed integuments, which are infiltrated with waxes and oils. In effect these waterproofing agents enable the seed coat to inhibit water absorption by the embryo. This results in a type of seed dormancy very characteristic of legume crops (clover and alfalfa). In the cocklebur (*Xanthium strumarium*), seed-coat-induced dormancy is imposed by oxygen impermeability. The environment itself can break this type of seed-coat dormancy through alternating temperature extremes of freezing and thawing. The extreme heat induced by forest fires is especially effective.

Fungal and bacterial infection of the seed coat can also lessen seed-coat-induced dormancy by reducing impermeability to oxygen, carbon dioxide, and water. Seed-coat dormancy can be artificially overcome through scarification, a process in which physical abrasion of the seed coat is accomplished through mechanical (sand or carborundum powder) or chemical (sulfuric acid) means.

Seed-coat-induced dormancy can also result from mechanical resistance due to extremely hard, rigid integuments commonly found in conifer seeds and other tree species with hard nuts (for example, macadamias and filberts). This dormancy may persist even though the seed coat may be hydrated. Once the seed coat dries out and is subsequently rehydrated, germination will usually follow. Other mechanisms previously described can also alleviate this type of seed-coat-induced dormancy.

2. *Embryonic factors.* The morphological state of the embryo is yet another physical factor affecting dormancy. Often the embryo is in a rudimentary stage when the seed is shed from the maternal plant, as exemplified by members of the orchid family and certain deciduous trees, such as ginkgo and European ash. Dormancy will usually cease in these plants as the embryos reach an adequate state of maturation.

Embryos can be physiologically mature, yet still remain dormant within the seed due to the lack of a metabolite essential for growth. These embryos are referred to as physiologically immature, and examples exist throughout the Rosaceae family and temperate-zone woody species. Dormancy in this instance can be broken by exposure to temperatures just above freezing or through commercially applied stratification. Such processes permit afterripening of the embryo, which leads to subsequent germination.

Endogenous regulation of dormancy. Studies dealing with dormancy have resulted in searches for endogenous plant hormones which regulate the process. In fact, the discovery of several plant growth regulators evolved from the close scrutiny of endogenous compounds derived from dormant structures.

Dormancy inducers. Early studies involving dormant buds of ash (*Fraxinus americana*) and birch (*Betula pubescens*) revealed the presence of high concentrations of a growth inhibitor or dormancy-inducing and -maintaining compound. This compound was later identified as abscisic acid. As buds of these trees began to grow and elongate, the levels of abscisic acid fell appreciably, supporting the role for abscisic acid in the regulation of dormancy.

Abscisic acid is also important in the regulation of seed dormancy, as exemplified by seeds of ash in which abscisic acid levels are high during the phase of growth inhibition, but then decline rapidly during stratification, resulting in germination. *See* ABSCISIC ACID.

Dormancy-releasing agents. Evidence from buds and seeds of other plant species, including apple (*Malus*), black currant (*Ribes nigrum*), and willow (*Salix fragilis*), show that abscisic acid levels drop sharply as dormancy is broken. In conjunction with decreased levels of abscisic acid, the endogenous supply of many growth promoters, such as gibberellins, cytokinins, and auxins, have been reported to rise during budbreak in sycamore (*Acer pseudoplatanus*) as well as in Douglas fir (*Pseudotsuga menziesii*). Levels of these dormancy-releasing compounds also correlate well with the breaking of seed dormancy. Dormant filbert seeds (*Corylus avellana*) usually require a period of chilling to break dormancy. Endogenous levels of gibberellic acid in these seeds show a marked increase in direct proportion to the degree of chilling. In fact, exogenous application of gibberellic acid can substitute for the chilling requirement. *See* AUXIN; CYTOKININS; GIBBERELLIN.

Cytokinins (substituted adenyl compounds) and auxins (for example, indole acetic acid, or IAA) are rather dualistic in nature, making it difficult to categorize them either as dormancy inducers or inhibitors. Exogenous cytokinins, for example, can overcome the inhibition of axillary bud dormancy. The breaking of bud dormancy in aspen (*Populus*) coincides with an increase in cytokinin activity. Although cytokinins are usually regarded as dormancy inhibitors, they can also induce tuberization which

eventually results in the production of a dormant structure (potato tuber). Auxins are also difficult to evaluate in terms of dormancy since they retard growth in several cases (for example, apical dominance in cereal grasses and in conifers) and promote it in others (tissue culture of embryos from various plant species). *See* APICAL DOMINANCE.

Exogenous regulation of dormancy. In addition to endogenous hormones, there are a variety of compounds that can break dormancy in plant species when they are applied exogenously. Many of these substances are synthetic derivatives or analogs of naturally occurring, dormancy-releasing agents. For example, the synthetic cytokinin 6-benzyladenine can stimulate growth of quiescent citrus (*Citrus aurantifolia*) buds by means of external application, and foliar sprays of growth inhibitors, such as the sterols paclobutrazol and fluriprimidol, can advance vegetative and floral budbreak in the peach (*Prunus persica*). The effect of the application of exogenous hormones is, however, very dependent on the genotype.

Synthetic chemicals can also be used in conjunction with naturally occurring compounds to enhance or break dormancy. A combination of 6-benzyladenine and the gibberellins GA_4 and GA_7 is effective in releasing dormant apple buds, whereas either chemical alone is relatively ineffective. Such chemical treatments are important in enhancing budbreak that can be greatly reduced when mild winters result in a lack of adequate chilling to trigger the production of endogenous dormancy-releasing hormones.

Mineral salts can also be beneficial in partially substituting for the natural chilling requirement to break dormancy. This rest substitution effect has been noted for foliar applications of potassium nitrate and thiourea, which selectively force apple flower and vegetative buds, respectively.

Interaction of dormancy-related compounds. The hormonal regulation of dormancy can best be perceived as a balance between dormancy inducers or maintainers and dormancy-releasing agents. This has already been discussed in studies involving fluctuating levels of abscisic and gibberellic acids in buds from several tree species. The same situation prevails in seeds where growth is closely regulated by the balance between different hormones and not necessarily by their absolute concentrations.

In pea (*Pisum sativum*) growth of the seed correlates well with hormonal fluctuations of abscisic, gibberellic, and indole acetic acids (**Fig. 1**). In this instance gibberellic acid and indole acetic acid levels rise with an increase in fresh weight and nutrient uptake by the embryo. The decline in seed growth coincides with a sharp rise in abscisic acid content and concomitant drop in gibberellic and indole acetic acids. *See* PLANT HORMONES.

Environmental versus internal factors. The physical environment exerts a marked influence on dormancy. The plant, however, needs a receptor system to perceive changes in the environment so it can translate them into physiological responses which in most cases are under hormonal control. In the

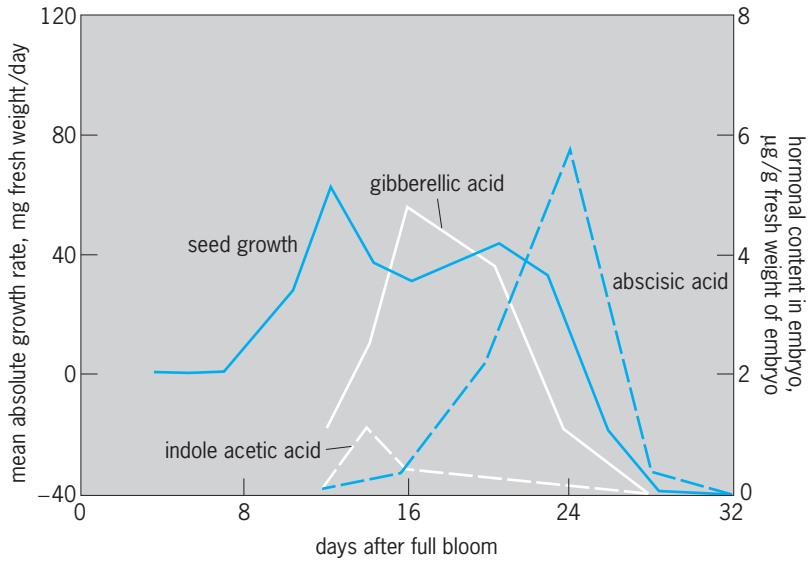

Fig. 1. Seed growth and hormonal levels in an embryo of *Pisum sativum*. Relative to indole acetic acid content, abscisic acid and gibberellic acid levels are 5 and 100 times less, respectively. (*After C. J. Eeuwens and W. W. Schwabe, Seed and pod wall development in Pisum sativum in relation to extracted and applied hormones, J. Exp. Bot., 26:1–14, 1975*)

case of changing day length or photoperiod, phytochrome may serve as a receptor pigment. Phytochrome essentially favors the production of either abscisic acid (short days) or gibberellic acid (long days; **Fig. 2**). Stress conditions, such as limited water or nutrient availability, favor the production of

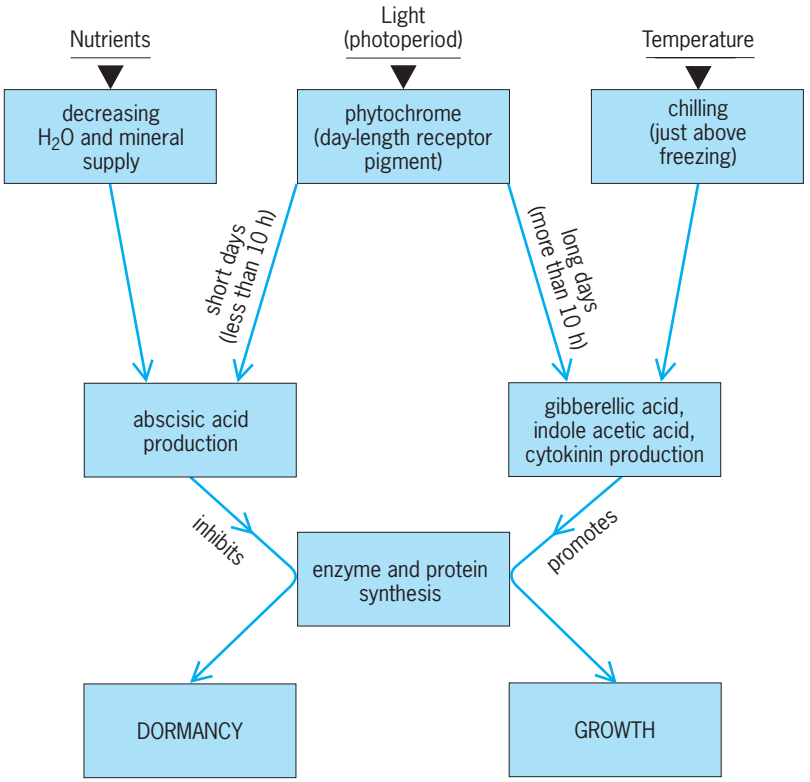

Fig. 2. Generalized scheme depicting the influence of the environment on endogenous pools of dormancy inducers and maintainers (abscisic acid) and dormancy releasers (gibberellic acid, indole acetic acid, and cytokinins), which in turn regulate dormancy via an effect on enzyme and protein synthesis in general.

abscisic acid, whereas a period of chilling often promotes synthesis of gibberellic acid and other compounds generally considered as growth promoters. The hypothesis of a receptor system such as phytochrome controlling the balance between levels of abscisic acid and gibberellic acid is intriguing, especially since the synthesis of these physiologically interactive hormones is probably associated with a common metabolic pathway originating with mevalonic acid. In this manner a definitive link between the environment and physiological changes (mediated via pools of endogenous hormones) could be established. *See* PHYTOCHROME.

Such endogenous pools are environmentally influenced and could in turn control dormancy physiologically. However, the mode of action of these endogenous growth regulators can only be postulated at this time. Whatever the specific mechanism, it probably involves the regulation of gene action at the level of deoxyribonucleic acid (DNA) and ribonucleic acid (RNA), which subsequently controls protein synthesis. In this framework, abscisic acid is believed to repress the functioning of nucleic acids responsible for triggering enzyme and protein synthesis needed for growth (Fig. 2). Gibberellic acid, on the other hand, promotes synthesis of enzymes essential for germination as in the case of α-amylase production that is crucial for barley seed growth. *See* BUD; NUCLEIC ACID; PLANT GROWTH; SEED.

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Dormouse

The name applied to the 31 species of Old World rodents of the family Gliridae. These animals are intermediate between squirrels and rats; they are vegetarians and arboreal like the former, but most have a general appearance like the latter.

The fat dormouse (*Glis glis*; see **illus.**), the largest member of the family, is found throughout Europe and western Asia. It resembles the squirrel because of its bushy tail, grows to a length of 8 in. (20 cm), including a 3-in. (7.5-cm) tail, and like other members of the family has 20 teeth, with a dental formula

The fat dormouse, *Glis glis.*

I 1/1 C 0/0 Pm 1/1 M 3/3. The fat dormouse prefers deciduous or mixed forests as a habitat; it hides in the forest during the day and forages during the night for fruits. It eats large quantities of food all summer and grows fat in preparation for winter, when it hibernates and sleeps until spring. These animals were bred in captivity by the Romans as a food delicacy.

The only species of dormouse native to Britain is the common dormouse (*Muscardinus avellanarius*), which is about the size of a house mouse. Its front legs are much shorter than the hindlegs, the thumbs are vestigial, and the big toes are reduced in size or rudimentary. Another species, found in central and southern Europe, is the garden dormouse (*Eliomys quercinus*), which is smaller than the fat dormouse but has similar habits. Both species are still eaten by humans in many places. *See* RODENTIA.

Charles B. Curtin

Dorylaimida

An order of nematodes in which the labia are generally well developed; however, many taxa exhibit a smoothly rounded anterior. The labial region is often set off from the general body contour by a constriction. The cephalic sensilla are all located on the labial region. When there is no constriction, the labial region is defined as that region anterior to the amphids. The amphidial pouch is shaped like an inverted stirrup, and the aperture is ellipsoidal or a transverse slit. The stoma is armed with a movable mural tooth or a hollow axial spear. The anterior portion of the tooth or spear is produced by a special cell in the anterior esophagus. The esophagus is divided into a slender, muscular anterior region and an elongated or pyriform glandular/muscular posterior

region. There are generally five esophageal glands with orifices posterior to the nerve ring. In some taxa there are three glands, and in others seven have been reported. The esophagointestinal valve is well developed. The mesenteron is often clearly divided into an anterior intestine and a prerectum. Females have one or two reflexed ovaries; when there is only one, the vulva may shift anteriorly. Males have paired equal spicules that are rarely accompanied by a gubernaculum. The males often have the ventromedial preanal supplements preceded by paired adanal supplements.

There are seven dorylaimid superfamilies: Actinolaimoidea, Belondiroidea, Diphtherophoroidea, Dorylaimoidea, Encholaimoidea, Nygolaimoidea, and Trichodoroidea.

Actinolaimoidea contains some of the most distinctive species with remarkable elaborations of the stoma which, however, always bears the characteristic axial spear. Feeding habits among the group vary from algae feeding to predation.

Belondiroidea is a diverse superfamily whose principal common characteristic is a thick sheath of presumed spiral muscle around the basal swollen portion of the esophagus. Nothing is known about the feeding habits of this group of nematodes, frequently encountered in soil near plant roots.

Diphtherophoroidea is presumed ancestral to the plant-parasitic superfamily Trichodoroidea. All known forms are soil inhabitants normally found in association with plant roots.

Dorylaimoidea includes representatives (family Longidoridae) that are important plant parasites and vectors of plant viruses throughout the world. The remaining families are algal feeders in soil and freshwater environments.

Encholaimoidea is a superfamily that is commonly found in soil near plant roots.

Nygolaimoidea is a superfamily of predaceous nematodes. The nygolaimoids have probably had the most varied taxonomic history among the Dorylaimida, having been treated as genera in Dorylaimoidea, a family in Leptonchoidea, and a superfamily in Dorylaimina.

The plant-parasitic Trichodoridea causes a condition called stubby root, in which root growth is arrested with little or no noticeable necrosis. Some members of this superfamily are known vectors of plant viruses. *See* NEMATA.

Armand R. Maggenti

Dosimeter

A device used to measure the dose of ionizing radiation received by an individual. There are many types of dosimeters with varying characteristics and capabilities appropriate for different applications. They differ in sensitivity, energy range, and species of radiation to which they respond. Some can be read out directly by the wearer, while others must be sent to a specially equipped facility to determine the dose.

Air ionization chamber dosimeter. This dosimeter is usually about the size and shape of a fountain pen with a clip suitable for carrying in a pocket, and is sometimes referred to as a pen meter or pocket dosimeter. The outside housing is usually made of plastic. Inside, a conductive cylinder and an insulated central electrode form an air-filled capacitor which is charged by connecting to a charging voltage supply. X-rays or gamma rays absorbed in the air space ionize the air, discharging the capacitor in proportion to the dose received. *See* CAPACITOR; GAMMA-RAY DETECTORS; X-RAYS.

The dose is read by connecting the dosimeter to a quartz fiber or other type of electrometer which reads the charge remaining on the capacitor. Some dosimeters contain an electrometer which can be viewed using a built-in lens and calibrated scale. Some also contain a built-in charging device. *See* ELECTROMETER.

The useful dose range is about 5–300 millirad (50–3000 microgray), depending on the capacitance. Due to insulator leakage, these dosimeters discharge slowly, a few millirads per week or more. *See* IONIZATION CHAMBER.

Film badge dosimeter. The badge housing is of plastic or metal, typically 50 mm square (2 in.) by 12 mm (0.5 in.) thick and equipped with a clip for attaching to clothing. It contains one or more film packets, usually standard dental x-ray film packets. The film is developed in a standardized procedure, and the optical density is measured and correlated to exposure, greater exposure being darker. *See* RADIOGRAPHY.

The side of the badge housing is equipped with one or more filters of lead, aluminum, copper, silver, or cadmium which cover different parts of the film and leave part uncovered. These filters modify the response of different areas of the film and thereby provide more information to correct the response to approximate that of tissue. For example, the uncovered portion responds down to low-energy x-rays and moderate-energy beta particles, while the lead filter area responds only to the more deeply penetrating higher energies. The high thermal neutron cross section of cadmium can be used to enhance the thermal neutron response via the gamma ray emitted on neutron capture. *See* BETA PARTICLES; NEURON; THERMAL NEUTRONS.

The effective dose range for x-rays and gamma rays is from about 30 mrad (300 μGy) to thousands of rads (tens of grays). Film badge dosimeters are fairly stable when stored under cool, dry conditions, with lifetimes suitable for readout at monthly but not longer intervals.

Thermoluminescent dosimeter. The physical packaging can be similar to a film badge with filters incorporated, but the film is replaced by pieces of thermolumiscent materials. Absorbed radiation energy excites electrons of the thermoluminescent material, creating excited metastable states. These states remain excited until the material is subsequently heated to a sufficient temperature, when they

deexcite, emitting visible light. The intensity of the emitted light is proportional to the absorbed radiation energy. The exposed material is placed in an automated instrument which heats it in a controlled manner, and the light is measured with a photomultiplier tube and the resulting signal is converted to a dose. If the exposed material is heated sufficiently to anneal all excited states, the dosimeter can be reused. *See* PHOTOMULTIPLIER.

Thermoluminescence has been observed in many materials, but four are commonly used in dosimeters. These are lithium fluoride (LiF), lithium borate ($Li_2B_4O_7$), calcium fluoride (CaF_2), and calcium sulfate ($CaSO_4$), usually activated with dopants such as manganese (Mn), titanium (Ti), and dysprosium (Dy), which enhance the light output. The light intensity versus temperature response of the material is termed its glow curve. To be useful, the glow curve must have a peak at a temperature sufficiently above ambient and storage temperatures to avoid fading, and also low enough to be conveniently read out in the reader. The details of the glow curve depend on the impurities and the fabrication history of the material.

Lithium fluoride is extensively used for dosimeters, activated with manganese and titanium. It has an average atomic number close to that of soft tissue, so it has tissue equivalent dose response without correction for x-rays and gamma rays.

Because of the high thermal neutron capture cross section of lithium-6 (^{6}Li), lithium fluoride also responds to thermal neutrons. By using a combination of ^{6}LiF and ^{7}LiF, the neutron contribution can be separated by subtracting one reading from the other. This response is strongly skewed toward low energy. To get a reading of neutrons incident upon and reflected from a person, an albedo configuration can be used. In one type, a three-layer sandwich is made with a ^{6}LiF/^{7}LiF pair on each side of a layer of thermal neutron absorber such as cadmium- or boron-loaded plastic. It is worn with one side close against the body. The side away from the body sees only incident thermal neutrons, while the body side sees primarily neutrons (of various energies) reflected and moderated by the body. The albedo response is obtained by subtracting a correction from the incident side. *See* THERMOLUMINESCENCE.

Electronic dosimeter. Usually rectangular in shape, this dosimeter ranges in size from a transistor radio to a pocket pager, though some devices resemble a fountain pen. It contains as sensing element a Geiger-Müller tube along with high-voltage supply, counting and control electronics, and a digital display, all powered by a battery. Battery life depends on features and usage but can be up to a year. *See* GEIGER-MÜLLER COUNTER.

A wide variety of features are available. For example, an audible dose rate indication can be provided by a chirp whenever a preset increment of dose is accumulated. A dose alarm can sound when a preset accumulated dose is reached. A dose-rate alarm can sound when a preset dose rate is exceeded, warning the wearer that he or she has wandered into a radiation area or a radiation source has been turned on. An electrical connection can be made with a computer for dose recordkeeping.

A less common feature is two-way wireless telemetry, permitting the dosimeter to be read and thresholds set remotely. A possible use of this capability is for a battlefield commander to assess the condition of his troops when tactical nuclear weapons have been employed.

Track etch dosimeter. The active element is a piece of CR-39 plastic (from Columbia resin 39), made by polymerization of allyl diglycol carbonate. When a high-energy charged particle, such as a proton recoiling from a fast neutron, traverses the material, it gives up energy and leaves a trail of chemical damage. The material is subsequently exposed to a suitable etchant such as sodium hydroxide (NaOH), and damage sites are preferentially etched, leaving pits which can be observed and counted under a microscope.

This dosimeter is sensitive only to high-energy neutrons, so it is usually used in conjunction with other dosimeters; for example, it is incorporated into a badge-type housing with thermoluminescent x- and gamma-ray and albedo dosimeters. The useful dose range is down to about 30 millirem (300 microsieverts). *See* PARTICLE TRACK ETCHING.

Bubble detector dosimeter. The operation of this dosimeter is based on the change to gas bubbles of superheated liquid (at a temperature above its boiling point) triggered by fast neutron interactions. A commercial bubble detector personal neutron dosimeter consists of a clear plastic tube with an aluminum fitting at one end, about the size of a large pen. It holds about 8 cm^3 (0.5 $in.^3$) of clear, solid elastic polymer containing about 10,000 superheated liquid droplets. On exposure to fast neutrons, small uniform-sized bubbles form. The bubbles are counted visually and a fast neutron dose calculated using a calibration factor determined for that dosimeter.

The dosimeter can be reset by screwing in a knob on the metal fitting which applies about 1.4 megapascals: (200 $lbf/in.^2$) pressure to the polymer, raising the boiling point and reliquefying the bubbles. Unscrewing the knob then relieves the pressure so that the droplets are again superheated, ready for reuse as a dosimeter. The dosimeter is temperature-sensitive and is available with a compensation feature if it needs to be used over varying temperatures. The useful dose range is down to less than 1 mrem (10 μSv), lower than any other fast neutron personal dosimeter. *See* HEALTH PHYSICS.

Peter Ryge

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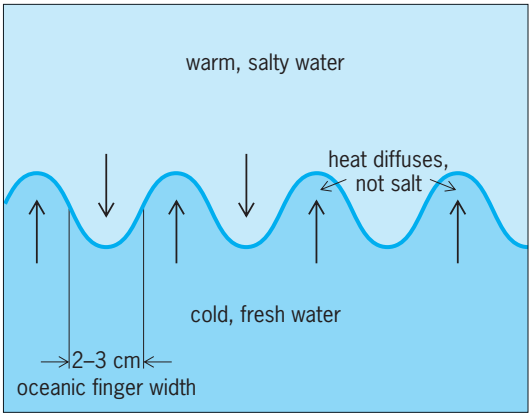
Double diffusion

A type of convective transport in fluids that depends on the difference in diffusion rates of at least two density-affecting components. This phenomenon was discovered in 1960 in an oceanographic context, where the two components are heat and dissolved salts. (Double diffusion can also occur in other fluid media, such as magmas, stellar and planetary interiors, and liquid metals.) Besides different diffusivities, it is necessary to have an unstable or top-heavy distribution of one component.

In the oceanographic context, if the unstable component is the slower-diffusing (salt), with the overall gravitational stability maintained by the faster-diffusing component (heat), then "salt fingers" will form. Since warm, salty tropical waters generally overlie colder, fresher waters from polar regions, this is a very common stratification in the mid- to low-latitude ocean. Salt fingers arise spontaneously when small parcels of warm, salty water are displaced into the underlying cold, fresh water. Thermal conduction then removes the temperature difference much quicker than salt diffusion can take effect. The resulting cold, salty water parcel continues to sink because of its greater density. Conversely, a parcel of cold, fresh water displaced upward gains heat but not salt, becoming buoyant and continuing to rise (see **illus.**). The fully developed flow has intermingled columns of up- and downgoing fluid, with lateral exchange of heat but not salt, carrying advective vertical fluxes of salt and to a lesser extent heat. In the ocean, the width of individual salt fingers is 2–3 cm, the distance over which thermal diffusion is effective. Their vertical extent can vary from a fraction of a meter to several meters, often occurring within interfaces separating well-mixed layers of 5–50-m vertical extent. When a series of layers and interfaces is found, it is termed a thermohaline staircase, because of the steppy appearance of the temperature and salt profiles. Staircase structures are maintained by the falling out of salt, which is made possible by the greater lateral diffusion of heat. They are quite common in certain oceanic regions, such as the western tropical North Atlantic and the Mediterranean Sea.



Initial growth of salt fingers arises when parcels of warm, salty water intrude into the cold, fresh water below. Heat escapes from these parcels at a greater rate than salt, because of its much larger diffusivity. The resulting cold, salty water is dense and thus continues to sink. Similarly, cold, fresh water displaced upward gains heat but not salt and becomes buoyant. The resulting network of up- and downgoing columns of fluid exchange heat laterally and carry salt downward. Such salt fingers are a few centimeters wide in the ocean, representing the scale over which thermal conduction can effectively release the gravitational potential energy stored in the unstable salt distribution.

The other form of double-diffusive convection occurs when the faster-diffusing component has an unstable distribution. In the ocean, this happens when cold, fresh water sits above warmer, saltier and denser water. Such stratifications are common in polar regions and in local areas above hot springs at the bottom of the deep sea. In this case, the greater diffusion of heat through a thin, high-gradient interface drives convection in adjacent mixed layers. Relatively little diffusion of salt occurs so that the effect is to transport heat but not salt vertically. Thermohaline staircases are formed in this instance as well, with the net heat flux through the system depending on the sharpness of the interfacial gradients. Because of the importance of diffusion across the thin interface, this type of double diffusion is termed diffusive convection. Both salt fingers and diffusive convection lead to separate vertical transport rates for heat and salt, and a flux of density which is up-gradient. Up-gradient density fluxes are necessary to maintain the thermohaline staircases; the action of turbulence or ordinary diffusion would give a down-gradient density transport and the smoothing away of steppy structure.

Recent measurements indicate that potentially disruptive turbulence is rare in the ocean interior, and the appropriate heat and salt stratifications quite common, making double diffusion a factor in many regions of the world ocean. The importance of double diffusion lies in its ability to affect water mass structure with its differential transport rates for heat and salt. This is believed to play a significant role in producing the "central water" masses of the subtropical gyres with well-defined relationships between temperature and salinity. *See* OCEAN CIRCULATION; OCEANOGRAPHY; SEAWATER. Raymond W. Schmitt

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Douglas-fir

A large coniferous tree, *Pseudotsuga menziesii* (formerly *P. taxifolia*), known also as red fir, belonging to the pine family (Pinaceae). It is one of the most widespread and most valuable tree species of western North America and ranks among the world's

Cone on a branch of Douglas-fir (*Pseudotsuga menziesii*).

most important. In the United States, this species is first in total stand volume, lumber production, and veneer for plywood. It is the most common Christmas tree in the West and is the state tree of Oregon. The name honors David Douglas, the Scottish botanical explorer who introduced the seeds to Europe. *See* PINALES.

Douglas-fir is a large to very large evergreen tree with narrow pointed crown of slightly drooping branches; at maturity it reaches a height of 80–200 ft (24–60 m) and a trunk diameter of 2–5 ft (0.6–1.5 m). In rare instances heights of 300 ft (90 m) are reported. The bark is dark or reddish brown, very thick, deeply furrowed, and often corky. The needlelike leaves spreading mostly in two rows are 0.75–1.25 in. (2–3 cm) long, flat, and flexible, with a very short leafstalk. Buds are distinctive, conical pointed, scaly, and dark red. The cones are elliptical, 2–3.5 in. (5–9 cm) long and light brown, with many thin rounded cone-scales, each above a longer distinctive three-pointed bract (see **illus.**). Seeds are paired and long-winged.

The natural distribution of Douglas-fir extends from southwestern Canada (British Columbia and Alberta) south through the western United States (Washington, Oregon, the Sierra Nevada, and the Rocky Mountains), and south to central Mexico. This species has been introduced into the eastern United States, Europe, and elsewhere.

Two varieties are distinguished: coast Douglas-fir of the Pacific region and Rocky Mountain or inland Douglas-fir. The latter has shorter cones with bracts bent backward. Though not so large, it is hardier and grows better in the East.

The trees form vast forests of pure even-aged stands, also with other conifers, growing rapidly and often attaining an age of 500 years or more. The largest commercial forests are in western Washington and western Oregon.

The wood is reddish, sometimes yellowish brown. Varying in weight and strength, it is moderately lightweight and moderately resistant to decay. Principal uses are building construction as lumber, timbers, piling, and plywood. The lumber is manufactured into millwork, railroad cars, boxes and crates, flooring, furniture, ships and boats, ladders, and so forth. Other products are railroad crossties, cooperage, mine timbers, fencing, and pulpwood. Douglas-fir is planted also for ornament, shade, shelterbelts, and Christmas trees.

Bigcone Douglas-fir (*P. macrocarpa*), the second native species, is local in mountains of southern California. It is a medium-sized tree with larger cones 4–6 in. (10–15 cm) long. Four or five other species are found in China and Japan. *See* FOREST AND FORESTRY; TREE.

Elbert L. Little, Jr.

Down syndrome

The set of physical, mental, and functional abnormalities that result from trisomy 21, the presence in the genome of three, rather than the normal two, chromosomes 21. Down syndrome is one of the most frequent genetic causes of mental retardation. The physical abnormalities that together give rise to the distinctive facial appearance associated with this condition include upslanting eyes with epicanthic folds (folds of skin over the inner corners), flatness of the bridge of the nose and of the middle portion of the face, and a tendency to protrude the tongue, especially when very young. Many other minor, functionally inconsequential abnormalities of the ears, hands, and feet may also be present, and stature is generally reduced.

Approximately 40% of affected individuals are born with congenital heart disease, usually an atrioventricular canal or ventricular septal defect, and obstruction of the intestinal tract (duodenal atresia or stenosis) also occasionally occurs during development. Both the heart and the intestinal anomalies are amenable to surgical correction. Only about a third of embryos and fetuses are actually liveborn. Individuals with Down syndrome are more than normally susceptible to infections, probably because of abnormalities of the lymphocyte system that result from abnormal development of the thymus, and hypothyroidism is relatively common. They also have an increased, although still low, likelihood of developing childhood leukemia. Despite these predispositions, their mean life expectancy, if they survive the first year of life, is now estimated to be between 55 and 60 years. *See* CONGENITAL ANOMALIES.

Nervous system abnormalities. Down syndrome is characterized by three salient clinical abnormalities of the nervous system. The first, recognizable immediately after birth and prominent in the first years of life, is a significant degree of hypotonia that produces a sense of floppiness or looseness. The second is mental retardation, which is generally moderate but may vary from mild to severe. The rate of acquisition of skills and capabilities is often normal during the first few months of life. However, there is then a decrease in rate so that the development of more advanced physical and intellectual skills, such as the ability to construct sentences, to walk stairs alone, and to dress, may be delayed by a year or more. Persons with Down syndrome have learned complex intellectual tasks such as reading, writing,

and arithmetic calculations, but this does not occur in all instances. The acquisition of speech appears to be more affected than other areas of cognitive development. *See* COGNITION.

There is evidence that the intellectual achievements of affected individuals reared in enriched environments, such as the home, are greater than was previously observed in individuals confined to institutions for the mentally retarded. However, even with children with Down syndrome reared at home, there is a progressive and virtually linear decline in developmental quotient (DQ) or intelligence quotient (IQ) starting within the first year. The same appears to be true for social quotients (SQ). Although a wide variety of treatments and therapeutic agents have been tried, including megavitamins and various drugs, there is not at present any accepted pharmacological treatment for Down syndrome. *See* INTELLIGENCE; MENTAL RETARDATION.

The third clinical abnormality of the nervous system, recognizable in 25 to 50% of individuals with Down syndrome living into the fifth decade, is the progressive loss of mental function resulting from the superimposition of a dementing process on the preexisting mental retardation. Associated with the development of these clinical features of dementia has been the appearance, in virtually all persons with Down syndrome older than 35 to 40 years of age, of the pathological and neurochemical changes indistinguishable from those found in the brains of individuals with Alzheimer's disease. Persons with Down syndrome thus constitute a unique population at very high risk of developing Alzheimer's disease, both pathologically and, in many instances, clinically. However, the nature of the association between the two conditions is unknown. Although Alzheimer's disease may be a direct consequence of genetic imbalance produced by trisomy 21, with increased expression of *APP* (the gene for the amyloid precursor protein), it is also possible that the trisomic condition serves to make individuals with Down syndrome more susceptible to those factors that lead to the development of Alzheimer's disease in a chromosomally normal aging population. *See* ALZHEIMER'S DISEASE.

Maternal age. The risk of having a child with Down syndrome increases with maternal age. The distribution of maternal age in the population of women having children is, therefore, the primary determinant of the overall incidence of Down syndrome. The overall incidence of Down syndrome in the newborn population is about 1 per 1000, and the maternal age-specific rate per 1000 live births increases from 0.6 at 20 years of age to 1 at 30 years, 2.6 at 35 years, 9.1 at 40 years, and 41.2 at 46 years. Analyses using deoxyribonucleic acid (DNA) markers have shown that nondisjunction, the failure of paired chromosomes 21 to separate properly during meiosis in the egg, accounts for 92% of all cases of trisomy 21 and is associated with either a decreased or an increased rate of meiotic recombination. The reason for the marked increase in nondisjunction with maternal age is not known, and there is no means for preventing it. *See* MEIOSIS.

Prenatal diagnosis. A reduction in the incidence of Down syndrome has been brought about by the use of prenatal diagnosis and selective termination of affected pregnancies. Prenatal diagnosis is performed by either chorionic villus sampling (obtaining a small piece of placental tissue) at the end of the first trimester or, more commonly, by amniocentesis (withdrawing amniotic fluid) in the second trimester. In either instance, the complement of chromosomes in the fetal cells is analyzed to determine whether there are any alterations in chromosome number or structure. Because these invasive forms of prenatal diagnosis are generally carried out in women 35 years and older, maternal serum screening has been introduced to identify pregnancies at increased risk, irrespective of the mother's age. This form of screening is based on the detection of altered levels of proteins and hormones in the maternal serum, and it must be followed up by amniocentesis to establish a definite diagnosis. Examination of the fetus by ultrasound is also being used as a means of screening pregnancies, but it is not diagnostic. *See* PREGNANCY; PRENATAL DIAGNOSIS.

Trisomy 21 and phenotype. The relationship between the presence of an extra chromosome 21 and the several phenotypic features of Down syndrome is unknown. Virtually complete sequencing of human chromosome 21 has revealed that there are about 240 known and predicted genes on the long arm (21q). Cases in which only part of chromosome 21 is triplicated have been intensively studied in an attempt to correlate particular phenotypic features of Down syndrome with specific regions or loci on the chromosome. Many of these features appear to be associated with imbalance in the so-called Down syndrome region surrounding and distal to the DNA marker *D21S55*. However, there is also evidence for contributions of genes outside this region, especially in the proximal part of 21q, and the region responsible for congenital heart disease seems to be confined to the distal part of 21q. Impaired cognition appears to be the result of the imbalance of genes located in several regions of chromosome 21. *See* CHROMOSOME; CHROMOSOME ABERRATION; GENE; HUMAN GENETICS.

Gene dosage effects. The immediate consequence of having an extra chromosome is a gene dosage effect (that is, an increase in expression of genes located on chromosome 21) for each of the loci present on this chromosome, and such gene dosage effects have been reported for several chromosome 21 loci. With possibly a few exceptions, the measured increases in activities or concentrations of products coded for by chromosome 21 in trisomic cells are close to the theoretically expected value of 1.5. However, changes in the synthesis of proteins coded for by genes elsewhere in the genome have also been observed, presumably as the result of genomic dysregulation resulting from the increased expression of chromosome 21 genes. This dysregulation is limited and does

not affect the vast majority of expressed genes. On the basis of studies in transgenic animal models, decreased platelet serotonin uptake and prostaglandin synthesis and abnormal myoneural junctions, all of which occur in Down syndrome, have been ascribed to increased activity of the enzyme CuZn-superoxide dismutase (SOD1). With this exception, and possibly the relationship between *APP* and Alzheimer's disease, it has not been possible to relate any component of the Down syndrome phenotype to overexpression of specific loci. To permit this problem to be attacked systematically, mouse models of Down syndrome have been made. *See* GENE ACTION.

Mouse models. It has been shown, by comparative gene mapping and genomic studies, that orthologs (similar genes in different species) of about half of the genes on human chromosome 21q are present on the distal part of mouse chromosome 16. The genes distal to *MX1* on human chromosome 21 are represented on mouse chromosomes 10 and 17. On the basis of these findings, several strains of mice segmentally trisomic for portions of the region of mouse chromosome 16 that is orthologous to human 21q have been generated to serve as genetic models of Down syndrome. These animals, which reproduce several of the phenotypic features of Down syndrome, including the craniofacial dysmorphology, male sterility, delayed cognitive maturation, impaired learning, and structural alterations of the cerebellum and certain neurons, are being used to investigate the contributions of specific genes to the phenotype of Down syndrome. Charles J. Epstein

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Drafting

The making of drawings of objects, structures, or systems that have been visualized by engineers, scientists, or others. Such drawings may be executed in the following ways: manually with drawing instruments and other aids such as templates and appliqués, freehand with pencil on paper, or with automated devices, described below.

Drafting is done by persons with varied backgrounds. Engineers often draft their own designs to determine whether they are workable, structurally sound, and economical. However, much routine drafting is done under the supervision of engineers by technicians specifically trained as drafters. *See* COMPUTER GRAPHICS; DESCRIPTIVE GEOMETRY; ENGINEERING DRAWING. Charles J. Baer

Templates. As the complexity of designs has increased and the use of standard parts has become widespread, drawings have become increasingly stylized. Graphic symbols have replaced pictorial representations. The next step in the evolution of drafting techniques has been the introduction of templates that carry frequently used symbols, from which the draftsman quickly traces the symbols in the required positions on the drawing (**Fig. 1**).

Automated drafting. Where the design procedures from which drawings are developed are repetitive, computers can be programmed to perform the design and to produce their outputs as instructions to automatic drafting equipment. Essentially, automated drafting is a method for creating an engineering drawing or similar document consisting of line delineation either in combination with, or expressed entirely by, alphanumeric characters.

The computer receives as input a comparatively simplified definition of the product design in a form that establishes a mathematical or digital definition of the object to be described graphically. The computer then applies programmed computations, standards, and formatting to direct the graphics-producing device. This method provides for close-tolerance accuracy of delineation and produces at speeds much greater than possible by manual drafting. In addition, the computer can be programmed to check the design information for accuracy, completeness, and continuity during the processing cycle.

Transmittal of the design definition information from the engineer to the computer is in algebraic terms that can be understood and processed by the computer program. This may be accomplished in two ways: (1) by the engineer trained in computer-understandable language, or (2) an intermediary who translates the engineer's notes and sketches into computer language for processing by the computer to drive a graphics output device.

In increasing use is the cathode-ray tube (CRT) console on-line with the computer to provide direct

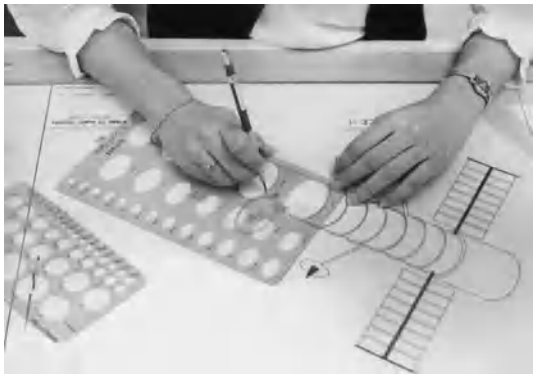

Fig. 1. Use of template in drafting.

interaction between a designer and a computer to develop the definition of a design (**Fig. 2**). The console contains a CRT on which the design is displayed as it is developed by the engineer writing with an electronic light pen to create line delineation and position characters; control keys to initiate the action to be performed on the CRT; function controls to call out program-stored information; and a keyboard, or typewriter, to apply the alphanumeric information required.

The engineering designer uses the light pen in a freehand mode to sketch the drawing and to position lines and program-stored symbols and characters. After a point is located or digitized by the light-sensing pen (which is a digitizer), the designer may type in a code for a symbol on the keyboard or may touch a symbol on the menu with the pen. A menu is a coded chart that shows graphic symbols. The symbols are displayed on the face of the CRT and are moved in sequential position by direction of the light pen. When the required symbols are in place, interconnecting lines are applied by direction with the light pen. During the development of the drawing, graphical images can be repositioned, or erased, instantaneously. The designer can observe the drawing as it progresses and call on the computer to provide computational assistance in developing the design. The information displayed on the CRT is stored in the computer memory, available to drive a graphics output device to produce a conventional engineering drawing when the design is complete (**Fig. 3**).

Graphics-producing devices include XY coordinate plotters, CRT plotters, and photocomposing units, each driven either on-line or off-line by a computer direction. XY coordinate plotters are available in models that produce highly precise drawings rapidly. Alphanumerics and symbols are either drawn by movement of the beam and drawing head, or the equipment may include a character-generator unit to imprint characters or symbols in a single stroke at the position directed by the computer.

CRT plotters produce drawings by exposing an image on sensitized film, usually 35 mm or smaller. By computer direction, the drawing image is displayed on a precision CRT face and recorded on film by a microfilm camera. There is also available a CRT plotter that produces full-size drawings up to 40 × 60 in. (100 × 150 cm) by sweeping the CRT head several times across the surface of a large sheet of sensitized film.

Photocomposing units are used primarily to produce drawings consisting of straight lines, symbols, and alphanumeric characters. Through a method of digitizing that identifies characters, lines, and their position in relation to each other, the drawing information is supplied to the computer where it is formatted to drive the photocomposing unit. The photo unit then projects a light beam successively through masks of the required symbols onto sensitized film.

The computer program required for processing design information to drive the graphics devices is unique for each type of equipment. Through these automated drafting methods, corrections and revisions can be made by providing the computer only

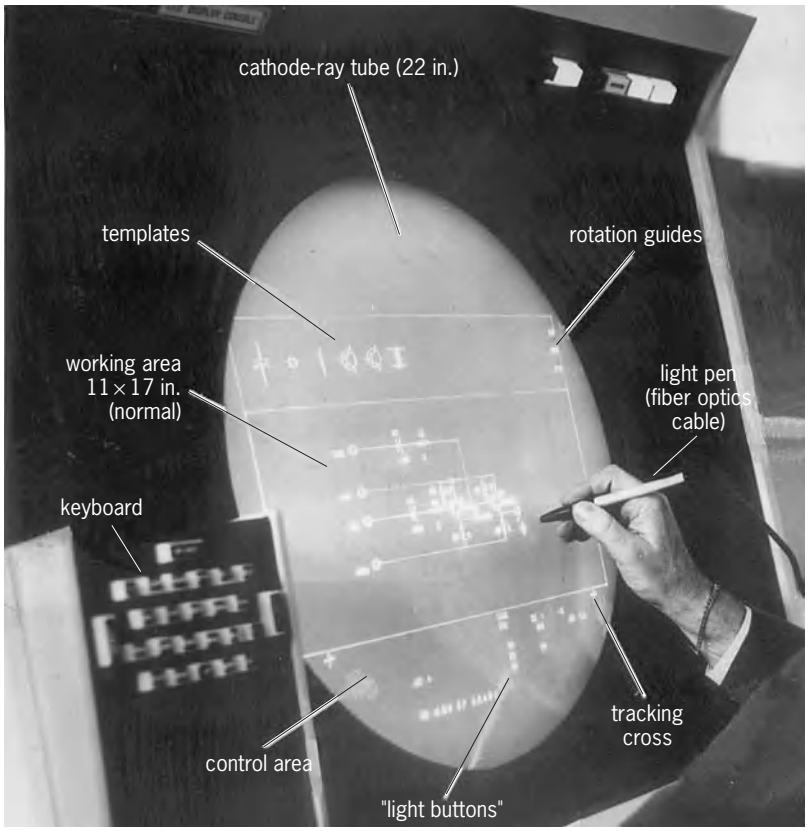

Fig. 2. Electronic light pen is used by designer on a cathode-ray tube to instruct computer in the drafting of a schematic. 1 in. = 2.54 cm. (*Lockheed*)

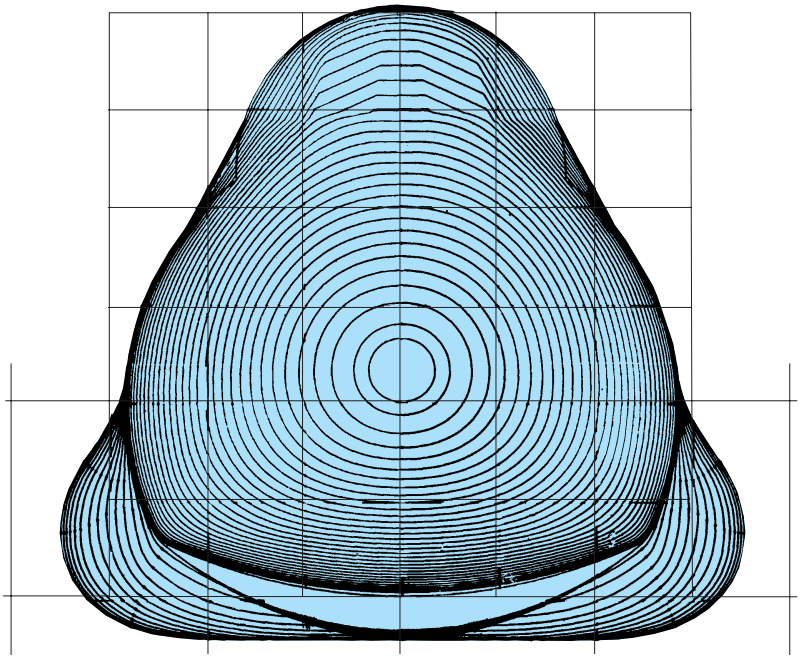

Fig. 3. Loft drawing of body stack for airplane fuselage was produced from design data by equipment such as that shown in Fig. 2. (*Lockheed*)

with the change information to revise the instructions for driving the output device. Tram C. Pritchard

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Drawing of metal

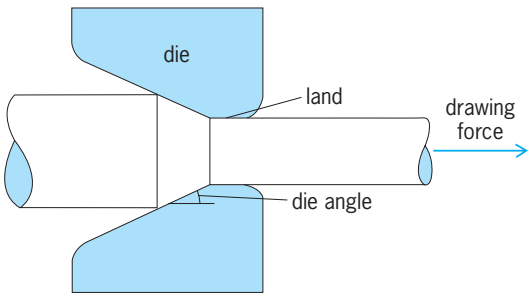
An operation wherein the workpiece is pulled through a die, resulting in a reduction in outside dimensions. This article deals only with bar and wire drawing and tube drawing. For deep drawing and other processes, as performed on sheet metal, *see* SHEET-METAL FORMING

Wire and bar drawing. Among the variables involved in the drawing of wires and bars are properties of the original material, percent reduction of cross-sectional area, die angle and geometry, speed of drawing, and lubrication. The operation usually consists of swaging the end of a round rod to reduce the cross-sectional area so that it can be fed into the die; the material is then pulled through the die at speeds as high as 8000 ft/min. Short lengths are drawn on a draw bench, while long lengths (coils) are drawn on bull blocks. Most wire drawing involves several dies in tandem to reduce the diameter to the desired dimension.

Die angles usually range from 6 to 15°, the actual angle depending on die and workpiece materials. Reductions in cross-sectional area vary 10–45% in one pass, although theoretically the maximum reduction per pass for a perfectly plastic material is 63%.

Die materials are usually alloy steels, carbides, and diamond. Diamond dies are used for drawing fine wires. The purpose of the die land is to maintain dimensional accuracy (see **illus.**).

The force required to pull the workpiece through the die is a function of the strength of the material, die angle, coefficient of friction, and reduction in cross-sectional area. The work applied to the process comprises three components: ideal work of deformation, friction work, and redundant work due to nonuniform deformation within the material. Depending on a number of factors, there is an optimum die angle for which the drawing force is a minimum. In cold-drawing, the strength of the material increases due to work hardening. The terminology for the increasing hardness is shown in the **table**.



Cross section of drawing die.

Hardness terminology for drawn wire

Terminology	Reduction in area in drawing, %
1/8 hard	11
1/4 hard	21
1/2 hard	37
Hard	60
Extra hard	75
Spring	84
Extra spring	90
Special spring	94

Temperature rise in drawing is important because of its effect on die life, lubrication, and residual stresses. Also, a defect in drawn rods is the rupturing of the core, called cuppy core. The tendency for such internal rupturing increases with increasing die angle, friction, and inclusions in the original material, and with decreasing reduction per pass.

The magnitude of residual stresses in a drawn material depends on the die geometry and reduction. The surface residual stresses are generally compressive for light reductions and tensile for intermediate or heavy reductions.

Extensive study has been made of lubrication in rod and wire drawing. The most common lubricants are various oils containing fatty or chlorinated additives, chemical compounds, soap solutions, and sulfate and oxalate coatings. The original rod to be drawn is usually surface-treated by pickling to remove scale, which can be abrasive and thus considerably reduce die life. For drawing of steel, chemically deposited copper coatings are also used. If the lubricant is applied to the wire surface, it is called dry drawings; if the dies and blocks are completely immersed in the lubricant, the process is called wet drawing. *See* LUBRICANT.

Tube drawing. Tubes are also drawn through dies to reduce the outside diameter and to control the wall thickness. The thickness can be reduced and the inside surface finish can be controlled by using an internal mandrel (plug). Various arrangements and techniques have been developed in drawing tubes of many materials and a variety of cross sections. Dies for tube drawing are made of essentially the same materials as those used in rod drawing.

Serope Kalpakjian

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Drier (paint)

A material added to paint formulations to facilitate the oxidation of oils. Linseed oil in its pure state normally requires about 3 days to dry to a hard film; linseed oil with driers added dries to a hard film in about 4 h.

The mechanism of drier reaction is not entirely clear. It is believed that the oil oxidation process produces peroxides that subsequently are destroyed, as radicals suitable for polymerization are formed, and that catalytic amounts of drier added to the oil accelerate peroxide destruction and the formation of free radicals. There is evidence that this mechanism may be altered when mixtures of driers are added to the oil, however, for mixtures accelerate the drying process faster than can any one drier alone.

Driers are salts of metals. Acids used to render metals soluble in oil, and therefore useful as driers, may be fatty acids, rosin, or naphthenic or octoic acids derived from petroleum. In commercially available driers, the percentages of salts dissolved in given amounts of mineral spirits are clearly defined so that additions can be accurately measured. Mineral spirit solutions containing mixtures of several metals are available when a special drying effect is desired.

Cobalt is the most reactive of drier metals, and normally is used in quantities amounting to less than 0.1% of the oil. Cobalt generally is regarded as a surface drier, and it is widely used as the only additive in thin-film paint formulations. In heavy-film paint formulations, another metal may be used to dry most of the film with cobalt as an auxiliary metal to dry the film surface and free it of tack.

Lead is less reactive than cobalt and is used in amounts of about 0.5% of the oil. Ecological requirements that paints be free from metallic contamination by lead may restrict the use of lead salts as a viable drier material, however.

Zirconium, in the form of zirconyl soap, has also been used as a commercial drier. Normal use ranges for zirconium as an auxiliary drier are between 0.1% and 0.4% in metal, based on vehicle solids. At these levels zirconium provides a thorough drying capability very close to that of lead. Its advantages are that it does not sulfide-stain, it is in the nontoxic category, and, because of excellent solubility, it does not impair gloss or gloss retention.

Manganese is somewhat less reactive than cobalt. It discolors certain oils, and for this reason is used chiefly in exterior paints subject to the bleaching action of air and sunlight.

Iron is not an effective drier at ambient temperatures, but it is often used in baking finishes. The color of iron salts usually prohibits their use in white paints.

Numerous other metals, including cerium and vanadium, have been used occasionally, and are effective driers. Their use normally is not economically justifiable, however. Zinc and calcium are employed only as auxiliary driers along with other metals.

Certain organic compounds such as 1,10-phenanthroline also catalyze the drying of oils and have been used for this purpose when freedom from all metallic contamination is required.

In general, small amounts of drier are essential to the formation of a satisfactory paint film within a reasonable time. Addition of large amounts of drier, however, can lead to premature embrittlement and failure of the paint film. *See* DRYING OIL; PAINT AND COATINGS.

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Drilling, geotechnical

The drilling of holes for gathering and evaluating earth materials in order to design and monitor construction projects. Geotechnical data are required for the accurate, safe, and efficient design and construction of buildings, bridges, highways, dams, and mining sites. The data are derived from analysis of soil and rock samples obtained by drilling. Soil sampling usually is done with a split-spoon sampler, a tube that opens lengthwise to remove the sample. Rock cores are obtained by diamond drilling, using a hollow, diamond-embedded drill bit to cut an intact rock sample. The samples are tested in laboratories for compressive and shear strength, grain size, weathering properties, moisture content, and consolidation projections. The results are used to determine the materials' supporting characteristics, ability to resist transmission of fluids as with dams, and ability to stand without lateral support as in highway cuts or quarries. In addition, tunneling and mining are accomplished safely only with an understanding of the properties of the soil and rock above and below the projected opening. In-hole or in-situ testing through boreholes drilled in a formation is done to determine the material's permeability, water yield, movement (with inclinometers), and settlement over time (with settlement indicators). *See* ENGINEERING GEOLOGY; ROCK MECHANICS; SOIL MECHANICS.

Elements of drilling. The drilling process requires that soil and rock particles be separated from the formation and removed from the hole. Penetration is achieved by crushing the rock as with a roller bit, by scraping, or by shearing the bottom of the hole with a drag or auger bit. Diamond drilling uses the extreme hardness of diamond to shear particles from the rock. Removal of particles from the hole is accomplished by mechanical means as with an auger, or with fluid (for example, water flush), airflow, or drill mud. In soft materials, a means (such as a steel casing) must be used to prevent the formation from collapsing into the hole.

All drilling requires the application of energy to the formation in the form of pressure, rotation, impact, or fluid flow. The action occurs by the use of tubular drill rods or augers with extensions.

Principal methods. Five commonly used drilling methods are (1) auger drilling of soils and soft rock using a helical screw with hardened bit, (2) rotary drilling with a roller bit and applied weight, (3) core drilling by rotary drilling in a circular groove or ring leaving a central core, (4) percussion drilling by the linear impact of the tool on the rock, and (5) sonic drilling, that is, fluidization of the formation by high-frequency vibration. Most often, a combination of

Fig. 1. Auger-core drill. (*Pennsylvania Drilling Co.*)

auger, rotary, and core drilling techniques is used for geotechnical drilling projects.

Auger drilling and soil sampling. Auger drilling is very common in the geotechnical soil sampling process (**Fig. 1**). Augers can have a solid stem with continuous helical flight (cutting edge) that conveys soils to the surface, or can have a hollow stem in combination with continuous flights enabling the taking of samples through the center of the augers (**Fig. 2**). Augers use hard, tungsten carbide bits to shear particles from the formation, which are conveyed to the surface by the auger flights that run to the top of the drill string. Augers are added usually in 5-ft (1.5-m) lengths, as needed.

Split-spoon samples are taken with the augers removed from the hole or through the hollow stem, continuously or at intervals. The use of a standard sampler and a 140-lb (64-kg) drive hammer is known as the standard penetration test. The sampler is driven into the ground by dropping the hammer 30 in. (760 mm) repeatedly and counting the blows to penetrate 6-in. (150-mm) intervals. This gives the geotechnical engineer an indication of the supporting characteristics of the soils. Samples from the inside of a split spoon are described and logged by the driller to give a continuous description of the formation drilled.

Rotary drilling. In this type of drilling, a rotary motion is induced in the drill rods by the drill rig and transmitted to the drill bit. The cutting action is accomplished by using a roller bit which has rolling cutters with hard projecting teeth or a drag bit which has fixed hardened cutting edges. The roller bit penetrates the rock by the crushing action of its teeth. Weight is placed on the bit by using heavy drill rods or collars or by pushing the rods from the surface with the drill rig while rotating the tools. The drag bit cuts with a scraping action at the bottom of the hole, lifting the rock particles away from the rock.

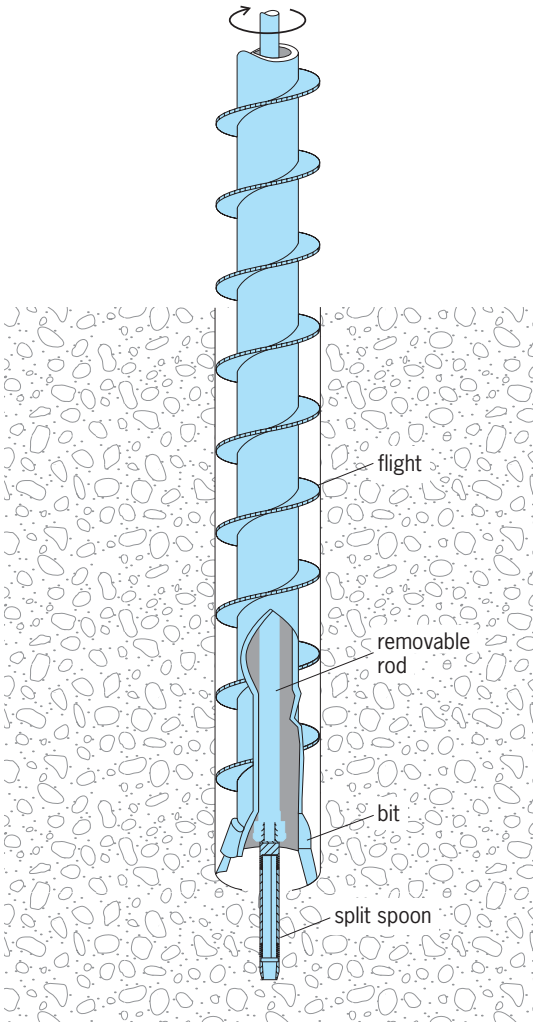

Fig. 2. Hollow-stem auger drill taking a soil sample.

Typical diamond bit core sizes.

Hole diameter, in. (mm)	Core diameter, in. (mm)
1.890 (48.0)	1.062 (27.0)
2.360 (60.0)	1.433 (36.4)
2.980 (75.7)	1.995 (50.6)
3.782 (96.0)	2.500 (63.5)
4.827 (122.6)	3.345 (85.0)

Removal of the cuttings is done by injecting air, fluid, or foam into the hole through a swivel that remains stationary as the drilling rods are rotated. The air or fluid is transmitted down the rods under pressure to the bit, where it cools the cutting surfaces and flushes the rock particles from the hole between the drill rods and the wall of the hole. Commonly, rotary drilling is used for drilling water wells using air as the flushing medium, and gas and oil wells using drilling mud. Drilling mud is a bentonite clay in combination with additives to change viscosity, density, and other properties. The mud has three functions in the deep gas and oil well drilling: it stabilizes the walls of the hole, cools the cutting edges, and flushes the cuttings from the hole.

Diamond drilling. A variation of rotary drilling is diamond drilling, which yields a continuous rock core for testing. This is also the main method for exploration of mineral deposits. In this type of drilling, the bit has diamond particles as the extremely hard cutting medium. Diamond bits are designed to cut an annular space and leave an intact rock core (see **table**). The core bit uses either a surface-set diamond pattern or an impregnated diamond matrix that wears away as it is used to expose new cutting surfaces. The surface set bits require larger diamonds (6–60 stones per carat weight) which are natural diamond stones. The impregnated diamond bits use synthetic diamonds that are very small (0.1 micrometer–1 millimeter). The core bits are used with core barrels that have an inner barrel liner, and a core lifter to enable the rock core to be pulled from the hole. For deeper holes, a retriever on a wire line cable is used to withdraw the inner barrel and core from the hole without pulling all the tools from the hole. *See* PROSPECTING.

Percussion drilling. A variation on rotary drilling is percussion drilling, where penetration is achieved by pneumatically powered impact. This method is often used for drilling blast holes for excavation as in highway cuts and quarries. The drill unit has a power unit similar to a rotating jackhammer that provides rapid impact blows to the drill string. Air is used to flush the rock fragments from the hole. Holes using this method are usually 4 in. (100 mm) in diameter or less.

Another percussion drilling method is downhole hammer drilling, where the impact action is generated at the bottom of the hole and not by the drill rig at the surface. The downhole hammer is powered by high-pressure air which actuates a piston in the hammer to act against the hammer bit. Rapid penetration rates can be achieved without using drill collars or using a heavy drill rig to push on the rods. The bits use tungsten carbide buttons (bumps) to fracture the formation.

Sonic drilling. This method is used mainly for environmental projects. The sonic drill has a special drill head that induces a powerful, high-frequency resonance in the drill string causing the formation to fluidize and allowing the tools to penetrate it. Energy is generated using rapidly rotating weights in the drill head to impart linear high-frequency vibrational energy to the drill string. Practical applications are for drilling of difficult formations such as glacial tills and boulders. Samples are taken using a dual-tube casing system, where the outer tube holds up the hole and the inner tube collects a continuous core.

Other drilling methods. Additional drilling methods are used to meet different geology and hole size and depth requirements.

One of the oldest methods is cable tool drilling, where a chisel-type bit is raised and lowered rapidly with a cable. Water is added, if needed, and the cuttings are extracted by lowering a tubular bailer and pulling out the water and rock or soil particles. This method is still commonly used for drilling water wells.

Auger mining uses the conveyor action of continuous-flight augers for actual mining operations in large-diameter horizontal holes. Drilling proceeds horizontally into a hillside or highwall of a coal mine to recover coal beyond the limits of strip mining.

Auger stem, drilled pier, and caisson drilling uses a short helix run on a solid or telescoping stem. The helix has a cutting bit to gouge out the formation and feed the material to the spiral helix. The bit is withdrawn and the soil and rock fragments are spun off by centrifugal force. This method is commonly used for building concrete piers to support buildings or other structures. It is suitable for holes 12–72 in. (304–1830 mm) or larger in diameter.

Raise drilling is used for constructing large airshafts and access ways for mines. The term "raise" comes from the drilling done from the bottom up. A small, usually less than 15-in. (380-mm) hole is drilled from the surface to the mine entry. A large-diameter bit is attached to drill rods at the bottom, and the bit is pulled to the surface while rotating. Cuttings are allowed to fall to the mine floor, where they are loaded on cars that are brought to the surface for disposal. Because of the tremendous forces that can be applied in the drilling by raise boring equipment, holes over 20 ft (6 m) in diameter can be drilled.

Reverse-circulation drilling is another variation on rotary drilling. Pressurized air or fluid is applied to the annulus between the drill rods and the hole wall and the cuttings come out through the center of the drill rods. Larger holes can be drilled with this method with less air supply. An air-assist line is usually used to help cuttings rapidly get to the surface.

Shaft drilling is shaft sinking by drilling a hole the size of the shaft, as contrasted with conventional shaft sinking where excavation is done by drilling and blasting. The smooth arch of the hole

circumference often makes lining of the shaft unnecessary in hard, massive rock. The drilling is done with larger versions of rotary drilling or reverse circulation drilling.

Directional drilling is a technique used to change the direction of a hole while drilling. This is used for hitting, from one location, multiple producing zones in offshore oil and gas drilling. Directional drilling is widely used in pipeline and utility construction to avoid interference with surface structures or lines. It has proven to be a good method for crossing under streams and rivers, avoiding damage to the channel and conflicts with recreation and commercial activity. Turbodrills are used in which the drill rods do not turn and a fluid-driven turbine motor at the end of the drill string turns a bit. A slightly offset rod behind the motor serves to push the hole in the desired direction. Complex electronics or sensors from the surface monitor the location and direction of the hole. *See* TURBODRILL.

Site evaluation. Test drilling to investigate the supporting characteristics of underlying soil and rock at proposed construction sites is done in a variety of ways. Most common is the American Society for Testing and Materials Standard Penetration Test for soil properties. In this test, a standard sampler is driven into the soil with a specific weight falling a set distance. Many other tests are available both in place and in the soils laboratory to measure moisture content, plasticity, grain size, shear strength, permeability, axial strength, deformation over time, and other properties. Some of these tests are performed on disturbed samples obtained with the Standard Penetration Test, and some are performed on larger samples obtained with test pits or undisturbed sampling with Shelby tubes. In Shelby tube sampling, a thin-walled metal tube with a formed, sharpened edge is pressed into the soil with the hydraulic feed apparatus on the drill rig.

Rock core drilling provides rock samples for visual examination and laboratory testing. The cores are tested in the laboratory for strength, permeability, freeze-thaw cycle deterioration, weathering characteristics, deformation under load over time, and other characteristics.

In-place borehole testing for engineering properties also includes pressure testing, permeability testing, pressure meter testing, pump testing for aquifer information, and cross-hole seismic tests.

Purchasers of contaminated real estate can incur tremendous cleanup costs. As a result, site evaluation for chemical properties of the underlying soil and rock has become the standard for property transfer because of the prevalence of subsurface contamination. Drilling and sampling are done, and monitoring wells of appropriate materials are installed to determine the extent, if any, of contamination generated by prior illegal dumping or substandard manufacturing processes.

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Drone

An aircraft or helicopter that is flown without an on-board human pilot. These vehicles are now customarily termed "unmanned aerial vehicles" (UAVs). Until the late twentieth century, they were typically called drones or sometimes remotely piloted vehicles (RPVs) when a human operator controlled the vehicle remotely. Although guided missiles and cruise missiles have the same root heritage in the "aerial torpedoes" of the 1910s, missiles are not considered to be UAVs. Crewless versions of nonpowered aerial vehicles such as gliders and balloons can perform some of the same missions as UAVs, but are not considered to be UAV-class vehicles. *See* MISSILE.

Types. There are three basic types of UAVs: the preprogrammed, the smart, and the remotely piloted UAV (**Fig. 1**).

A preprogrammed UAV (Fig. 1*a*) responds to an on-board timer or scheduler and has no sensor contact with the ground. The UAV follows a set routine of maneuvers, altitude changes, speed changes, and course changes that are programmed through an autopilot to the UAV's control surfaces and engine throttle. The UAV can be recovered by a parachute at the end of the mission.

A smart UAV (Fig. 1*b*) carries various sensors and is equipped with an on-board computer. The ability of a smart UAV to make decisions governing course and altitude changes is limited only by its computer and sensor capacity. For example, a smart UAV could probably take off on its own from a given airport, navigate a circuitous route, make decisions en route based on weather or enemy radar action, fly to a second airport, and make a safe landing. *See* GUIDANCE SYSTEMS; NAVIGATION.

The remotely piloted UAV (Fig. 1*c*), probably the most common type of UAV, is under the constant control of an operator or pilot through radio links. The pilot or pilots can be located on the ground, in other aircraft, or on ships. A remotely piloted UAV system includes the aerial vehicle, launch and recovery systems, and ground or airborne stations for controlling the vehicle's flight and processing information from the vehicle. Typical missions for remotely piloted UAVs include reconnaissance or surveillance, jamming enemy communications, serving as decoys, and acquiring or attacking enemy targets. A UAV can acquire and maintain a lock on an enemy target using a laser designator. A decoy provides a false electronic signature to make it appear like a crewed aircraft to enemy radar. Reconnaissance or surveillance UAVs can be equipped with radar, low-light-level television, or infrared sensors that allow over-the-horizon reconnaissance imagery to be transmitted to ground commanders in real time. Rather than performing a single mission, UAVs which perform a multitude of different missions during a flight are becoming more common. *See* INFRARED IMAGING DEVICES.

Evolution. The first military experiments with UAVs began during World War I with the U.S. Navy–Sperry flying bomb and the U.S. Army–Kettering Bug. These were simply biplanes typical of that

era, controlled by a preprogrammed autopilot system. The British also experimented with a variety of drones in the 1920s, including the Larynx and DeHavilland Queen Bee, which were remotely piloted through radio control. World War II saw the debut of the combat version of UAVs, called UCAVs, although they were referred to as drones at the time. In the Pacific theatre, the Navy used TDR-1 Assault Drones to drop 2000 lb (900 kg) of bombs on Japanese ground targets. While the TDR-1s conducted their attack through antiaircraft fire, they were controlled a safe distance away by a circling crewed aircraft through television and telemetry links. Sixty-two percent of the 50 drones successfully damaged their targets while none of the orbiting crewed aircraft were damaged. In the European theatre, Project Aphrodite and Anvil used crewless B-17s under remote control to attack hardened targets in occupied France and Germany. These aircraft were loaded with as much as 24,000 lb (11,000 kg) of explosives. Take-off for these large aircraft was more difficult to control so a two-person crew initially flew the plane from England and then bailed out as a trailing crewed aircraft took over the controls via television and telemetry links. The lessons learned from these World War II experiences were applied by the Navy to send radio-controlled crewless Grumman Hellcats off of aircraft carriers to attack North Korean targets during the Korean War.

The next generation of remotely piloted UAVs emerged during the 1960s. These high-performance, crewless aircraft were flown during the Vietnam War on surveillance and reconnaissance missions. Crewed C-130 aircraft were used to air-launch, control, and then recover the reconnaissance UAVs. More than 3400 missions were flown over Southeast Asia, primarily by modified Ryan BMQ-34 Firebees (**Fig. 2**). These missions produced significant intelligence and targeting information without risking aircrews, but use of the intelligence was delayed until the UAV could be recovered and the film could be processed and interpreted. Extensive use of the modified BQM-34s in Southeast Asia revealed not only the remotely piloted UAV's promise but also its overall operational reliability and low vulnerability in combat. The most frequently used version of the BQM-34 series was the AQM-34L (model 147SC). A total of 1651 SC UAVs was launched combat, and 87.2% returned.

During the 1970s and 1980s, crewless aircraft were also effectively used as countermeasures against air defense systems. The Israelis flew Northrop MQM-74 Chukar-IIs against Egyptian air defenses during the war in 1973. In 1982, the Israelis destroyed most of the Syrian air defense radar warning and tracking system in just 2 h. Small UAVs simulating crewed attack aircraft decoyed Syrian radars into revealing their locations. Still more unpiloted and remotely piloted UAVs swarmed in to overwhelm the Syrian air defense system with false targets. These precursor tactics allowed Israeli crewed aircraft to rapidly destroy Syrian radars, as well as surface-to-air missile sites and interceptor aircraft.

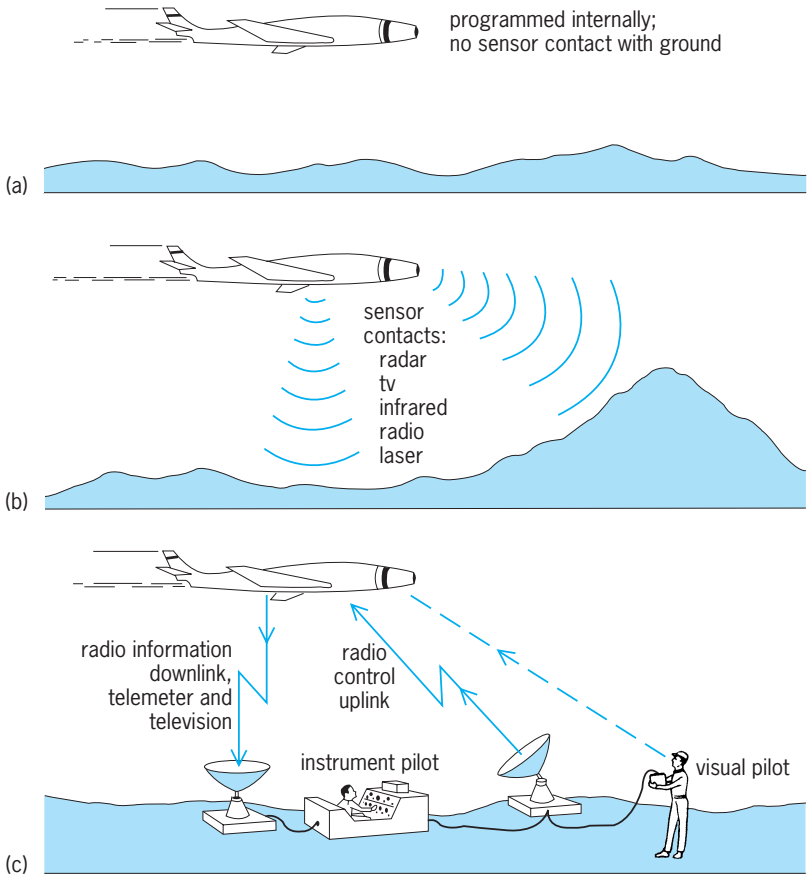

Fig. 1. UAV control schemes. (*a*) Preprogrammed UAV. (*b*) Smart UAV. (*c*) Remotely piloted UAV.

Fig. 2. Modified Ryan reconnaissance vehicles. (*a*) High-altitude version, with wing span of about 27 ft (8 m). It photographed from altitudes of 60,000 ft (18,000 m). (*b*) Low-altitude version with stretched fuselage, large wing tips, and end plates on horizontal stabilizer. It made photo runs at 1500 ft (450 m). (*Teledyne-Ryan*)

Fig. 3. General Atomics Predator UAV. (*General Atomics Aeronautical Systems, Inc.*)

Ship-borne Israeli Aircraft Industries (IAI) Pioneer remotely piloted UAVs were used as gunfire spotters during the Persian Gulf War of 1991. The Pioneers provided television imagery of the downrange effectiveness of naval gunfire to offshore U.S. battleships. The psychological effect of the heavy bombardment was so great that on one occasion, Iraqi soldiers surrendered to an orbiting Pioneer before naval gunfire commenced. UAVs such as the IAI/TRW Hunter and General Atomics Predator (**Fig. 3**) provided real-time reconnaissance imagery during the Balkans War of the mid- to late-1990s in Bosnia and Kosvo. However, there was a time delay between target identification by the UAV ground operator and a request for tactical air support up the chain of command. This led to the escape of several hundred Serbian vehicles before the arrival of attack aircraft. Based on this experience, some Predator UAVs were modified to carry AGM-114 Hellfire antitank missiles and a laser target designator. These "low-flying" modified Predator UAVs together with "high-flying" Northrop Grumman Global Hawk reconnaissance UAVs (**Fig. 4**) flying at altitudes up to 60,000 ft (18,000 m) have been used effectively in the Afghanistan and Iraq wars.

Fig. 4. Northrop Grumman Global Hawk. (*Northrop Grumman*)

Status. The U.S. Air Force and the U.S. Navy have a fly-off competition for a new-generation of UCAVs. The Boeing X-45 and the Northrop Grumman X-47 prototypes both have a stealthy design to evade enemy radar in order to attack high-value targets which may be too dangerous for crewed aircraft to attack. A capability to enable UCAVs to refuel in the air using crewed tanker aircraft is also being considered. Although large UAVs provide excellent real-time reconnaissance imagery, their relative high cost and small numbers make it difficult to provide this information to units smaller than the battalion level. In order to resolve this issue, lower-cost, less capable small UAVs with a size on the order of radio controlled model airplanes are being introduced in the U.S. Marine Corps. Even smaller Micro Air Vehicles (MAVs) are also undergoing current research and development.

About half the countries in the world operate UAVs of one form or another. The United States spends more than half of the world's annual outlays on UAVs and UAV-related research due to the large expense associated with highly capable military UAVs. However, more than half of the UAVs in the world are radio-controlled helicopter-type UAVs for agricultural purposes in Japan. The Federal Aviation Administration has begun to consider how to integrate UAVs into the civil airspace alongside private planes and passenger jets. As the performance and reliability of autopilots continue to improve, passenger-carrying aircraft may, themselves, someday become a form of a UAV by eliminating the on-board human pilot. *See* AUTOPILOT.

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Drought

A general term implying a deficiency of precipitation of sufficient magnitude to interfere with some phase of the economy. Agricultural drought, occurring when crops are threatened by lack of rain, is the most common. Hydrologic drought, when reservoirs are depleted, is another common form. The Palmer index has been used by agriculturalists to express the intensity of drought as a function of rainfall and hydrologic variables.

The meteorological causes of drought are usually associated with slow, prevailing, subsiding motions of air masses from continental source regions. These descending air motions, of the order of 600

or 900 ft/day (200 or 300 m/day), result in compressional warming of the air and therefore reduction in the relative humidity. Since the air usually starts out dry, and the relative humidity declines as the air descends, cloud formation is inhibited—or if clouds are formed, they are soon dissipated. In the United States, the area over which such subsidence prevails may involve several states, as in the 1977 drought over much of the Far West, the 1962–1966 Northeast drought, or the dust bowl of the 1930s over the Central Plains.

The atmospheric circulations which lead to this subsidence are the so-called centers of action, like the Bermuda High, which are linked to the planetary waves of the upper-level westerlies. If these centers are displaced from their normal positions or are abnormally developed, they frequently introduce anomalously moist or dry air masses into regions of the temperate latitudes. More important, these long waves interact with the cyclones along the Polar Front in such a way as to form and steer their course into or away from certain areas. In the areas relatively invulnerable to the cyclones, the air descends, and if this process repeats time after time, a deficiency of rainfall and drought may occur. In other areas where moist air is frequently forced to ascend, heavy rains occur. Therefore, drought in one area, say the northeastern United States, is usually associated with abundant precipitation elsewhere, like over the Central Plains. *See* CYCLONE.

After drought has been established in an area, there seems to be a tendency for it to persist and expand into adjacent areas. Although little is known about the physical mechanisms involved in this expansion and persistence, some circumstantial evidence suggests that numerous "feedback" processes are set in motion which aggravate the situation. Among these are large-scale interactions between ocean and atmosphere in which variations in ocean-surface temperature are produced by abnormal wind systems, and these in turn encourage further development of the same type of abnormal circulation. Atmospheric interactions called teleconnections can then establish drought-producing upper-level high-pressure areas over North America. Then again, if an area, such as the Central Plains, is subject to dryness and heat in spring, the parched soil appears to influence subsequent air circulations and rainfall in a drought-extending sense.

A theory supported by numerical modeling studies suggests that over very dry areas, particularly deserts, the loss of heat by radiation relative to the surroundings creates a slow descending motion. This results in compressional warming, lowered relative humidity in the descending air, and inhibition of rain. These processes are probably partly responsible for the tropical droughts observed in various parts of the Sahel (sub-Saharan) region during the 1970s. Of course, interactions with other parts of the atmosphere's prevailing circulation as indicated earlier also play a role in tropical droughts.

Finally, it should be pointed out that some of the most extensive droughts, like those of the 1930s dust bowl era, require compatibly placed centers of action over both the Atlantic and the Pacific. *See* CLIMATE HISTORY.

Another type of drought came to light in Indonesia and the surrounding area, sometimes extending as far westward as India. These droughts may last a couple of years and are associated with the wind systems which create El Niño, an abnormally warm mass of water near the Equator, which can have an east-west extent of hundreds, sometimes thousands, of miles. This water, as much as a few degrees above normal temperature, interacts with the overlying atmosphere to produce rising air motion with more clouds and rainfall in the eastern Pacific (from Peru westward) but sinking air motion with inhibition of rainfall over the western Pacific including Indonesia. While occurring on average every five or so years, El Niño is not periodic and is difficult to predict.

The Sahel (sub-Saharan) droughts are the most long-lasting, sometimes persisting for more than a decade. These droughts impact on agriculture and cattle production to the extent that hundreds of thousands of people starve to death in countries such as Ethiopia. These Sahel droughts are associated with abnormal displacements of the Intertropical Convergence Zone, a west-to-east-oriented area where the trade winds of both hemispheres converge. This is a zone of enhanced rainfall, so that its variable latitudinal position often determines how much rain occurs in parts of the sub-Saharan region. Once desiccation sets in, air-land feedbacks help to perpetuate the drought. *See* SAHEL; TROPICAL METEOROLOGY.

In view of the immense scale and complexity of drought-producing systems, it will be difficult to devise methods of eliminating or ameliorating them.

Jerome Namias

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Drug delivery systems

Systems that deliver controlled amounts of a therapeutic agent over a prolonged period or in a desirable temporal fashion, thereby maintaining required plasma or tissue drug levels. Controlled drug delivery systems can take a variety of forms, such as mechanical pumps, polymer matrices or microparticulates, and externally applied transdermal patches. The

development of systems requires the coordinated effort of engineers, physicians, surgeons, chemists, and biologists.

Controlled drug delivery has long been used to treat certain diseases. Individuals with serious bloodborne infections are treated by continuous intravenous infusion of antibiotics. By monitoring the antibiotic concentration in the individual's blood and adjusting the rate of infusion, the physician precisely controls the delivery of the agent. Because this treatment is expensive and requires hospitalization, its use is limited to acute, life-threatening illnesses. In addition, routes of drug administration other than intravenous infusion are frequently required. This drug delivery concept has been applied more widely in the development of a variety of novel strategies.

Pumps. The most obvious approach for controlled drug delivery involves miniaturization of the familiar infusion system. Mechanical pumps, either totally implantable or requiring percutaneous catheters, have been used to deliver insulin, anticoagulants, analgesics, and cancer chemotherapy. Although the technology is mature and is accepted by many individuals requiring treatment, this approach has certain disadvantages. Drugs are maintained in a liquid reservoir prior to delivery, so only agents that are stable in solution at body temperature can be used. The devices are bulky, expensive, and suitable only for highly motivated individuals who can visit their physicians regularly. Because pumps deliver precisely controlled amounts of therapeutic agents, they may find additional important applications, particularly when coupled with implanted biosensors (electronics designed to continuously measure the concentration of a disease-specific analyte in tissue) for feedback control. *See* BIOELECTRONICS.

Polymers. Many disadvantages of controlled drug delivery via pumps can be avoided by using controlled-release polymers as delivery vehicles. In the early 1960s, the first methods for releasing drugs from biocompatible polymers were developed, and many approaches to polymeric controlled release have been tested. As a result, techniques for the production of polymer implants that slowly release drug molecules have become available. Nondegradable polymers, biodegradable polymers, swellable polymers, and biopolymers have been used as the basis of controlled drug delivery systems. For a specific therapeutic agent, the rate, pattern, and duration of drug release can be modified by selecting the appropriate biocompatible polymer and method of device fabrication. *See* POLYMER.

Nondegradable polymers. The first controlled-release polymer systems were based on semipermeable silicone elastomers. Hollow silicone tubes filled with a liquid suspension of a drug are immersed in water, whereupon the drug is released from the system by diffusion through the silicone wall. This technology permits the controlled release of any agent that can dissolve and diffuse through either silicone or poly(ethylene-co-vinyl acetate), the two most commonly used nondegradable, biocompatible polymers. The best example of this technology is the use of cylindrical silicone-based subdermal implants that release contraceptive steroids at a constant rate for 5 years following implantation and provide economical, reliable, long-term protection against unwanted pregnancy. An intrauterine device with a controlled-release poly(ethylene-co-vinyl acetate) coating delivers contraceptive steroids directly to the reproductive tract; local release of certain agents can reduce the overall drug dose and, therefore, reduce or eliminate side effects produced by drug toxicity in other tissues. *See* BIRTH CONTROL.

Transdermal controlled-delivery devices are similar in design to implantable reservoir systems. A single flat polymer membrane is placed between a drug reservoir and the skin. The polymer membrane is designed to control the rate of drug permeation into the skin over a period of several hours to 1 day. Transdermal delivery systems have been produced for agents that can penetrate the skin easily, such as nitroglycerin, scopolamine, clonidine, estradiol, and nicotine.

Solid matrices of nondegradable polymers can also be used for long-term drug release. In comparison to reservoir systems, these devices are simpler (since they are homogeneous and, hence, easier to produce) and potentially safer (since a mechanical defect in a reservoir device can lead to dose dumping). However, it is more difficult to achieve constant rates of drug release with nondegradable matrix systems. Constant release can sometimes be achieved by adding rate-limiting membranes to homogeneous matrices, yielding devices in which a core of polymer-drug matrix serves as the reservoir. In other cases, water-soluble, cross-linked polymers can be used as matrices; release is then activated by the swelling of the polymer matrix in water.

Biodegradable polymers. When nondegradable polymers are used for implanted delivery systems, the polymer must be surgically removed at the end of therapy. This procedure can be eliminated by the use of biodegradable polymers, which dissolve following implantation. Biodegradable polymers, first developed as absorbable sutures in the 1960s, are usually employed as matrix devices. Many chemically distinct biodegradable polymers have been developed for use as drug delivery devices. These polymers must be biocompatible (that is, provoking no undesirable or harmful tissue response) and the degradation products must be nontoxic. To provide reproducible drug delivery, these polymers must also degrade in a controlled manner. Of the polymers tested, poly(lactide-co-glycolide) has been used most frequently. Most importantly, small microspheres of biodegradable polymers can be produced as injectable or ingestible delivery systems; injectable poly(lactide-co-glycolide) microspheres that release a peptide are used for the treatment of prostate cancer in humans. The chemistry of the biodegradable polymers can be exploited in other ways to achieve controlled delivery, for example, by linking drugs to the polymer matrix through degradable covalent bonds.

Proteins and other macromolecules. Recombinant proteins and polypeptides are produced in large

quantities by the biotechnology industry. Many of these molecules, such as the recombinant tissue plasminogen activator, human insulin, and human growth hormone, have important applications in human health, although they are generally less stable than conventional drugs and more difficult to deliver. Protein drugs are difficult to use in humans, because either they are eliminated very quickly from the body or are toxic when delivered systemically at the doses required to achieve a local effect. Controlled-release polymers may overcome these problems by slowly releasing the protein into the blood stream over a long period or by releasing protein into a local tissue site, thus sparing systemic exposure. *See* GENE ACTION; GROWTH FACTOR; HORMONE; PHARMACOLOGY; TOXICOLOGY.

W. Mark Saltzman

Liposomes. Since the mid-1970s, liposomes (artificial phospholipid vesicles) have been considered promising delivery systems for biologically active substances because they are biologically inert and biocompatible; very rarely produce toxic or antigenic reactions; protect the drugs enclosed within them from the destructive action of the external media by sterically preventing their contact with inactivating factors in the blood and tissues; can deliver their content inside cells and even inside different cell compartments; and water-soluble drugs can be captured by the inner aqueous space of liposomes, while lipophilic compounds can be incorporated into the liposomal membrane.

Many different methods have been suggested to prepare liposomes of different sizes (from 50 nm to more than 1 μm in diameter) and properties. The most frequently used techniques are ultrasonication (the application of high-frequency sound waves to a lipid suspension in water, resulting in the formation of closed vesicles), reverse phase evaporation (the vacuum evaporation of a lipid-containing water-organic mixture, leaving the liposomal suspension in the water), and detergent removal from mixed lipid-detergent micelles by dialysis or gel filtration (upon removal of the detergent, micelles spontaneously convert into bilayered structures, liposomes).

To increase liposome accumulation in the desired area, the use of targeted liposomes has been suggested. Liposomes with a specific affinity for an affected organ or tissue were believed to increase the efficacy of liposomal pharmaceutical agents. Immunoglobulins (antibodies), primarily of the IgG class, and their fragments (which are attached to the surface of the liposome) are the most promising and widely used targeting moieties for various drugs and drug carriers including liposomes, since they can specifically recognize and bind certain characteristic components of target organs and tissues. Different methods have also been suggested to achieve longer circulation times of liposomes in vivo, including coating the liposome surface with inert, biocompatible polymers, such as polyethylene glycol (PEG), which form a protective layer over the liposome surface and slows down liposome recognition by opsonins (substances in blood serum that accelerate phagocytosis of foreign cells to which antibodies are attached) and subsequent clearance. Liposomes with increased longevity and targetability have also been prepared by adding both the protecting polymer and the targeting moiety to the liposome surface. *See* ANTIBODY; IMMUNOGLOBULIN; OPSONIN.

Most liposomes are internalized by cells via endocytosis and destined to lysosomes for degradation. To achieve liposome-mediated drug delivery inside cells, particularly into a cytoplasmic compartment, pH-sensitive liposomes are frequently used. Lipids used for their preparation are able to undergo certain structural transitions at pH values slightly below normal physiological levels, enabling the destabilization of the liposomal and endosomal membranes and drug release.

The most successful application of liposomes for drug delivery is in cancer therapy, for example, clinical application of doxorubicin (a widely used anticancer drug for both solid tumors and blood or bone marrow tumors) trapped within long-circulating liposomes demonstrated decreased toxicity and increased tumor accumulation. Liposomes are also frequently used as pharmaceutical delivery systems for various protein and peptide drugs; deoxyribonucleic acid (DNA), for example, for gene therapy or enhanced production of therapeutic proteins (liposomes containing positively charged lipids are especially effective in binding DNA and delivering it into cells); and short interfering ribonucleic acid (siRNA) for inhibiting translation of the disease-causing protein. *See* CELL PERMEABILITY; LIPID; LIPOSOMES; PHOSPHOLIPID.

Micelles. Micelles are colloidal particles, often spherical, with the size in a nanometer range, into which many amphiphilic molecules (which have a hydrophilic, or water soluble, head attached to a long hydrophobic, or water insoluble, tail) self-assemble spontaneously. In an aqueous environment, hydrophobic fragments of amphiphilic molecules form the core of a micelle, which is segregated from the environment by hydrophilic parts of the molecules forming the micelle corona. Micelles possess a number of significant advantages as potential drug delivery systems for poorly soluble pharmaceuticals, since the hydrophobic core of micelles may be used as a cargo space for encapsulation of a variety of sparingly soluble therapeutic and diagnostic agents. Such encapsulation substantially increases the bioavailability of pharmaceuticals, protects them from destructive factors upon parenteral administration (via injection), and beneficially modifies their pharmacokinetics (the way that drugs move through the body after they are administered) and biodistribution.

Polymeric micelles prepared from certain amphiphilic co-polymers consisting of both hydrophilic and hydrophobic monomer units represent a separate class of pharmaceutical micelles. The majority of micellar preparations, especially those with anticancer drugs, are still at the preclinical development stage. *See* COLLOID; MICELLE; POLYMER.

Novel systems. There are a number of novel drug delivery systems at different stages of preclinical

development and early clinical trials. These systems include microemulsions, lipoprotein-based drug carriers, lipid nanocapsules, lipid-coated microbubbles, dendrimers (polymers having a well-defined branched structure), cells and cell ghosts (such as red blood cell ghosts, that is, cells in which part of the inner content is replaced with a drug), aerosols, magnetic nanoparticles that can be concentrated in the required region in the body under the action of an external magnetic field, and niosomes (liposome-like structures prepared of nonionic surfactants). Attempts have also been made to make drug delivery systems that are stimuli-sensitive, that is, capable of releasing DNA or an entrapped drug in response to certain physiological changes, such as decreased pH values and/or increased temperature in pathological areas. The application of nanosized drug delivery systems represents an important area of nanomedicine. *See* NANOTECHNOLOGY.

Vladimir P. Torchilin

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Drug resistance

The ability of an organism to resist the action of an inhibitory molecule or compound. Examples of drug resistance include disease-causing bacteria evading the activity of antibiotics, the human immunodeficiency virus resisting antiviral agents, and human cancer cells replicating despite the presence of chemotherapy agents. There are many ways in which cells or organisms become resistant to drugs, and some organisms have developed many resistance mechanisms, each specific to a different drug. Drug resistance is best understood as it applies to bacteria, and the increasing resistance of many common disease-causing bacteria to antibiotics is a global crisis.

Genetic basis. Some organisms or cells are innately or inherently resistant to the action of specific drugs. In other cases, the development of drug resistance involves a change in the genetic makeup of the organism. This change can be either a mutation in a chromosomal gene or the acquisition of new genetic material from another cell or the environment.

Mutation. Any mutation that results in resistance to an antibiotic will provide a selective advantage to the bacterial strain in the presence of the antibiotic. Susceptible organisms are inhibited, while resistant ones replicate to become the predominant population. Single mutations that result in drug resistance are usually specific to a single drug or class of drugs. In order to prevent the emergence of resistance of a bacterial pathogen to a single antibiotic, physicians may prescribe a combination of antibiotics that have different mechanisms of action. In this way, a single genetic mutation conferring resistance to one class of drugs does not result in a selective advantage, since the other antibiotic may still be effective against that bacterial strain. The likelihood of a single organism developing resistance to two unrelated drugs by mutational events is much lower than to only one drug.

Gene acquisition. Organisms may acquire deoxyribonucleic acid (DNA) that codes for drug resistance by a number of mechanisms. Transformation involves the uptake of DNA from the environment. Bacteria that are able to undergo transformation are said to be competent; common species of bacteria that share this characteristic include *Haemophilus*, *Streptococcus*, and *Neisseria* spp. Once DNA is taken up into the bacterial cell, it can recombine with the recipient organism's chromosomal DNA. This process plays a role in the development and spread of antibiotic resistance, which can occur both within and between species.

Transduction, another mechanism by which new DNA is acquired by bacteria, is mediated by viruses that infect bacteria (bacteriophages). Bacteriophages can integrate their DNA into the bacterial chromosome. During excision of viral DNA for packaging into new virions, the virus may package not only its own DNA but parts of the bacterial DNA as well. This genetic material is then carried to the next bacterium infected by the bacteriophage.

Conjugation is the most common mechanism of acquisition and spread of resistance genes among bacteria. This process, which requires cell-to-cell contact, involves direct transfer of DNA from the donor cell to a recipient cell. While conjugation can involve cell-to-cell transfer of chromosomal genes, bacterial resistance genes are more commonly transferred on nonchromosomal genetic elements known as plasmids or transposons. Plasmids are extrachromosomal, self-replicating genetic elements which exist independently within the bacterial cell. In contrast, transposons do not exist independently within the cell except when transferred; during transference they are incorporated into the chromosomal DNA of the cell or into the DNA of a plasmid. Transposons are sometimes referred to as jumping genes. *See* DEOXYRIBONUCLEIC ACID (DNA).

Plasmids and transposons play an extremely important role in allowing bacteria to adapt to environmental change by accommodating the efficient exchange of genetic information both within and between bacterial species. Plasmids and transposons that carry genes coding for resistance to antibiotics are sometimes called R-factors or R-determinants. A

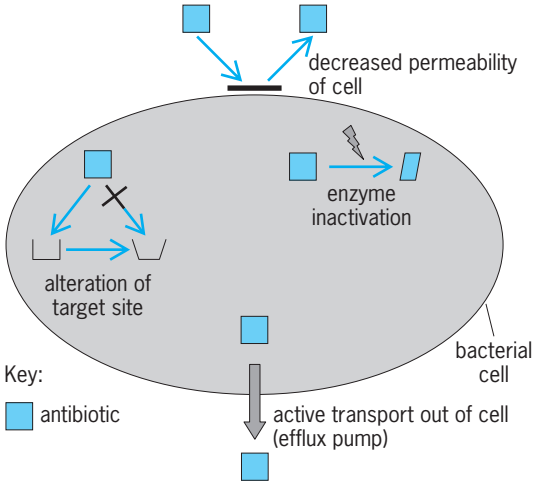

Fig. 1. Four common mechanisms of antibiotic resistance.

single plasmid or transposon often carries multiple resistance genes, and therefore provides a powerful means by which antibiotic resistance genes can disseminate. *See* PLASMID.

Mechanisms of resistance. The four most important antibiotic resistance mechanisms are alteration of the target site of the antibiotic, enzyme inactivation of the antibiotic, active transport of the antibiotic out of the bacterial cell, and decreased permeability of the bacterial cell wall to the antibiotic (**Fig. 1**).

Alteration of target site. By altering the target site to which an antibiotic must bind, an organism may decrease or eliminate the activity of the antibiotic. Alteration of the target site is the mechanism for one of the most problematic antibiotic resistances worldwide, methicillin resistance among *Staphylococcus aureus*. The most common human bacterial pathogen, *S. aureus* has acquired resistance to many different antibiotics. Against many hospital-acquired strains of methicillin-resistant *S. aureus*, only one currently approved antibiotic, vancomycin, remains active. If resistance to vancomycin becomes widespread among *S. aureus*, there exists the possibility of a postantibiotic era with respect to this pathogen, with no antibiotics available to effectively treat the bacteria.

Methicillin resistance among *Staphylococcus* species is mediated by a chromosomal gene (*mecA*) which codes for one of the penicillin-binding proteins (PBP2a) found in the cytoplasmic membrane of these bacteria. An alteration in this penicillin-binding protein decreases the affinity of the protein for all penicillin-related antibiotics. Since methicillin and related antibiotics cannot bind to this protein, they cannot interfere with cell wall synthesis. Methicillin-susceptible organisms do not carry the *mecA* gene.

Another problematic form of antibiotic resistance, penicillin resistance among *Streptococcus pneumoniae*, is also mediated by an alteration in the penicillin-binding proteins. There was a dramatic increase in resistance to penicillin among *S. pneumoniae* in the United States during the 1980s-1990s (**Fig. 2**). Furthermore, because all antibiotics that are based upon the penicillin ring structure (called the β-lactam ring) act by binding to penicillin-binding proteins, resistance to penicillin among *S. pneumoniae* is accompanied by a decrease in susceptibility to many other agents (such as other penicillins and cephalosporins). Since *S. pneumoniae* is the most common bacterial cause of pneumonia and meningitis in adults, this increasing resistance to penicillins and cephalosporins may have a tremendous impact on therapy for these serious infections.

Erythromycin resistance among some species of bacteria has also been found to be due to the alteration of a target site. In this case, a plasmid or transposon carries a gene encoding a ribosomal methylase. This enzyme causes a change in the ribosomal subunit that is targeted by erythromycin, thus causing decreased affinity and resistance to the drug. *See* BACTERIAL GENETICS.

Enzyme inactivation. The most common mechanism by which bacteria are resistant to antibiotics is by producing enzymes that inactivate the drugs. β-Lactam antibiotics (penicillins and cephalosporins) can be inactivated by enzymes known as β-lactamases. Hundreds of β-lactamases have been described; they can be both plasmid or chromosomally encoded, and have varying degrees of activity against the different β-lactam antibiotics. Many bacteria produce multiple β-lactamases. In response to the proliferation and spread of β-lactamases, the pharmaceutical industry has developed some β-lactam antibiotics that are more resistant to hydrolysis by these enzymes. In addition, some combination drugs have been produced which contain both a β-lactam antibiotic and a β-lactamase inhibitor; the inhibitor has high affinity for the β-lactamase enzyme, irreversibly binds to it, and thereby preserves the activity of the β-lactam antibiotic.

Bacterial resistance to aminoglycoside antibiotics is mediated by aminoglycoside-modifying enzymes. Acetyltransferases, phosphotransferases, and adenyltransferases have been described as modifying the structure of these drugs, rendering them less able to reach their target site. This is a common resistance mechanism employed by a broad range of bacterial species.

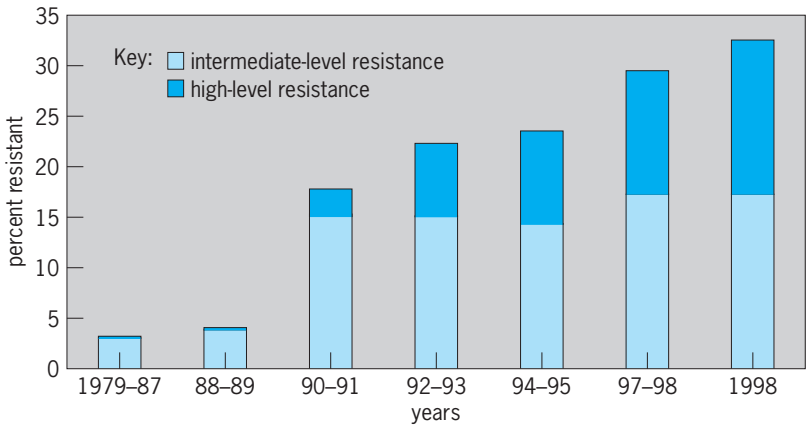

Fig. 2. Increasing penicillin-resistance rates among *Streptococcus pneumoniae* from United States medical centers since 1979.

Active transport of the drug. Active transport systems (efflux pumps) have been described for the removal of some antibiotics (such as tetracyclines, macrolides, and quinolones) from bacterial cells. In these situations, even though the drug can enter the bacterial cell, active efflux of the agent prevents it from accumulating and interfering with bacterial metabolism or replication.

Decreased permeability of the cell. Bacteria are intrinsically resistant to many drugs based solely on the fact that the drugs cannot penetrate the bacterial cell wall or cell membrane. In addition, bacteria can acquire resistance to a drug by an alteration in the porin proteins that form channels in the cell membrane. The resistance that *Pseudomonas aeruginosa* exhibits to a variety of penicillins and cephalosporins is mediated by an alteration in porin proteins.

Promoters. In the hospital environment, many factors combine to promote the development of drug resistance among bacteria. Increasing use of powerful new antibiotics gives selective advantage to the most resistant bacteria. In addition, advances in medical technology allow for the survival of sicker patients who undergo frequent invasive procedures. These factors allow bacteria to survive in sites to which antibiotic drugs do not penetrate well. When bacteria replicate in the presence of low levels of drug (levels too low to kill the bacteria or prevent replication), resistant mutants are more likely to emerge. Finally, poor infection control practices in hospitals allow for the unchecked spread of already resistant strains of bacteria. Some studies have demonstrated that only about half of physicians appropriately wash their hands before examining patients.

Outside the hospital environment, other important factors promote antibiotic resistance. The overuse of antibiotics in outpatient medicine and the use of antibiotics in agriculture exert selective pressure for the emergence of resistant bacterial strains. The spread of these resistant strains is facilitated by increasing numbers of children in close contact at day care centers, and by more national and international travel. *Streptococcus pneumoniae* develops resistance almost exclusively outside the hospital environment. Spread of penicillin-resistant strains of *S. pneumoniae* have been documented from one continent to another, and the highest prevalence of penicillin resistance in the United States was initially found in day care centers.

Impact. The most important impact of drug resistance is an increase in disease and death from infections due to drug-resistant strains of viruses or bacteria and from cancers due to chemotherapy-resistant neoplasms. The human immunodeficiency virus (HIV), which is estimated to have infected over 40 million people worldwide, has the ability to develop resistance to every antiviral agent yet approved for treatment. When HIV does become resistant to antiviral agents, treatment with these drugs predictably fails. *Staphylococcus aureus*, the most common human bacterial pathogen, is increasingly resistant to available antibiotics, and recently strains have been reported which are resistant to vancomycin, the last line of defense.

Increasing rates of drug resistance result in increased use of more expensive, and often more toxic, drugs. Prevention of transmission of resistant bacteria in hospitals requires more frequent isolation of patients, which is also expensive. Overall, the cost of antibiotic resistance has been estimated to exceed $30 billion annually in the United States.

Control. A multifaceted worldwide effort will be required to control drug resistance among disease-causing microorganisms. The first step is to perform good surveillance studies to define the extent of the problem. Ongoing programs to decrease the use of antibiotics, both in the clinics and in agriculture, will be necessary. The increased use of vaccines to prevent infection can help limit the need for antibiotics. Finally, the development of novel classes of antibiotics to fight emerging resistant bacteria will be required. Unfortunately, there is no guarantee that new antibiotics can be developed quickly enough to remain one step ahead of the most adaptable bacterial foes. *See* ANTIBIOTIC; BACTERIA. Daniel J. Diekema

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Dry-docking

The removal of a ship from its water environment, which needs to be done periodically to repaint the underwater portion of the hull, repair components exposed to the water, and conduct repairs and modernizations requiring cuts in the hull. At first, ships were just beached at low tide. This practice gave way to use of a marine railway where ships were hauled up the beach on rails.

The construction of the relatively low-capacity marine railways shifted to construction of high-capacity dry docks dug out of the ground (called graving docks). These "stone docks," which were constructed of granite stone, were large enough to accommodate the dry-docking of contemporary ships displacing 2000 long tons (2032 metric tons) and more in Europe and North America in the 1700s and 1800s. In the mid-1800s, the construction of these graving docks was followed by the construction of floating dry docks, which were also of relatively high capacity. These floating dry docks were portable and allowed ship repair to expand to the far reaches of the globe by the early 1900s.

As warship size increased and the shift from wood to steel occurred in the late 1800s, an ever-expanding complement of graving docks, floating docks, and marine railways emerged, especially during World War II, and this expansion continued until late in

the twentieth century. New methods of dry-docking ships emerged in the late twentieth century. These included heavylift ships, vertical lifts, and computer-controlled floating dry docks.

Reasons for dry-docking. Ships require dry-docking for many reasons. If antifouling paint has worn off from the ship's hull, the hull will foul with grass and other marine life such as barnacles, especially in warm salt water environments. The excessive accumulation of marine growth will eventually slow the ship down, causing it to be much less fuel efficient. Divers can clean the hull; however, the frequency of diver cleaning will continue to increase unless the antifouling coating is refurbished, which requires dry-docking. Many underwater repairs can also be accomplished by divers such as propeller replacement or placing on cofferdams for localized hull welding, painting, or sea valve replacement. When large-scale welding or overhhaul of numerous sea valves is required, again dry-docking is the most economical option. About every 8 to 12 years, a ship's propulsion shafting must be pulled and refurbished, requiring the ship to be in dry dock. Major repairs or modernizations below the water line of the hull that require large sections of the hull to be removed also require dry-docking. These include major machinery repairs or replacements where access cuts through the underwater hull are required.

Some ships are modified to increase their capacity. This is usually done by cutting the ship in half at midships, pulling the two halves apart forward and aft, and then inserting what is called a parallel midbody to lengthen the ship. Of course, this must be done in a dry dock.

Some large ships such as very large crude carriers (VLCC), enormous cruise ships, and aircraft carriers are built in relatively shallow graving docks called building docks. Ship erection units that had been constructed elsewhere in the shipyard are placed in the building dock and welded to the other units.

Some ships are constructed on what are called land-level transfer facilities. Large areas of a new construction shipyard are dedicated to such a facility. Ship units are erected on a carriage assembly and welded together. After the ship is assembled, outfitted, and painted, it is moved via the carriage on a rail track system onto an adjacent floating dry dock or vertical lift for launching.

The floating dry dock adjacent to a land-level transfer facility is typically landed on a concrete grid system so that a ship can be moved onto the dock via the carriage on rails. A rail bridge is placed across the short boundary between the dock and the land portion of the rail system. Once the ship is on the floating dry dock in the correct location, the floating dry dock is deballasted so that it floats off the concrete grid system. The floating dry dock has its own mooring system using a system of anchors, chains, and motor-driven winches to move itself over a dredged hole in the harbor (or river). The floating dry dock is then ballasted down and the ship floats off the blocks. The ship is then pulled out of the dry dock.

Frequency of dry-docking. The technology of underwater paint coating systems and increased ship husbandary (underwater ship cleaning) capabilities, maintained through diving operations, have allowed the intervals between dry-dockings to increase. The U.S. Navy has lengthened the period between dry-dockings from 3 years (around 1980) to as much as 12 years, depending on vessel class.

Care must be taken to properly examine shaft and strut bearings, shaft coating systems, and hull cathotic protection systems while a ship is in dry dock to assure that deficiencies are not later identified between these extended dry-docking periods that would necessitate unscheduled and usually very costly dry-dockings.

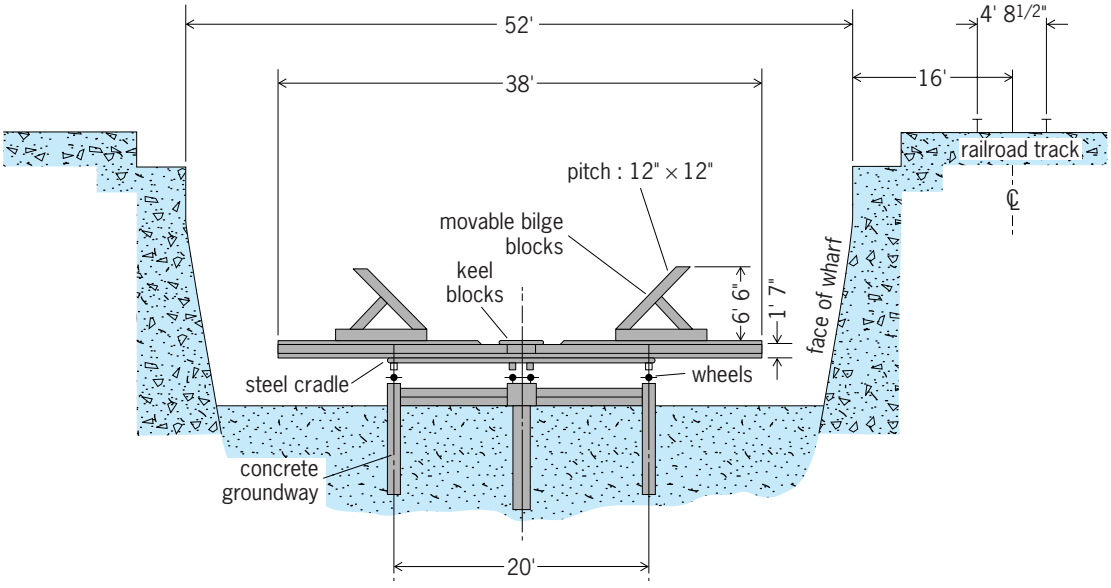

Fig. 1. End-view drawing (section looking inshore) of a marine railway circa 1955 at the Washington Navy Yard in Washington DC. This was the site of the first marine railway in United States. (*U.S. Navy*)

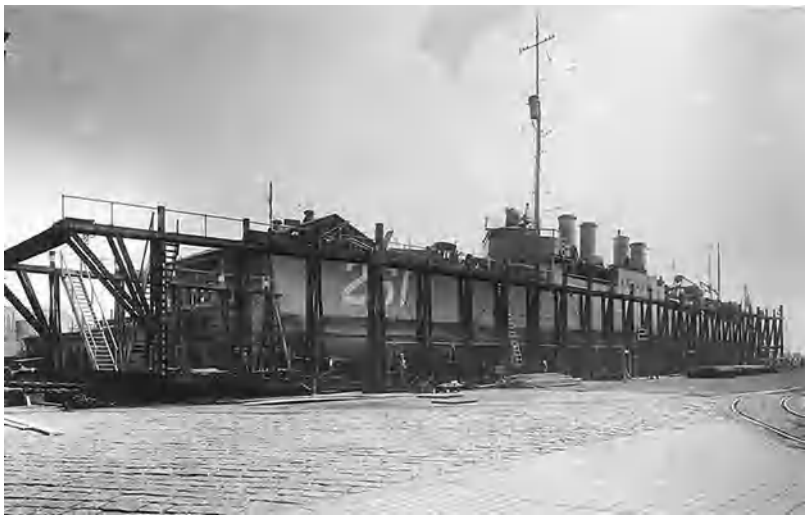

Fig. 2. USS *Delphy* (DD 261), a U.S. Navy World War I destroyer, on a marine railway after being hauled out of the water up the railway at the Boston Navy Yard (since closed) in 1919. (*U.S. Navy*)

Types of dry docks. The four types of dry docks are known as marine railways, floating dry docks, graving docks, and vertical lift docks. The size of the ship usually determines which type is used.

Marine railway. This consists of a cradle of wood or steel with rollers on which the ship may be hauled out of the water along a fixed inclined track leading up the bank of a body of water (**Figs. 1** and **2**). The advantages of a marine railway lie in the economy of the original construction and the relative low cost of maintenance. A marine railway is ideal for ships up to 5000 long tons (5080 metric tons).

Floating dry dock. This type of dry dock may be constructed of timber, steel, or concrete. Few timber floating dry docks remain. Some larger floating dry docks are built in sections called sectional floating docks which are secured together by steel plates (**Fig. 3**). These docks have the capability of self-docking their own sections for maintenance purposes. A floating dry dock is submerged enough to provide the required depth of water over the keel blocks, by filling its ballast tanks with water. The ship to be dry-docked is then positioned in the dock, the ballast tanks of the dock are pumped out by powerful electric pumps located within the dock's wingwalls, the ship lands on the keel blocks, the side blocks are hauled, and the ship is lifted out of the water.

The floating dry dock's ballast tanks are pumped out according to a carefully engineered plan. This plan is designed so that the dry-dock tank pumping sequence does not put undo stress on the dock structure and keeps the pontoon deck as straight as possible while the ship lands on the blocks and is raised out of the water.

The ship is positioned over the blocking system using lines or cables run from the dry-dock wing walls to the ship. Some of these lines accurately position the ship forward and aft, while others position the ship side to side. These lines also help move the ship in and out of the dry dock. Some ships have such deep projections below the keel (like propellers or transducers) that the ship's centerline must be kept to one side of the keel blocks while entering or leaving the dry dock.

Once the ship has fully landed on the keel blocks, the side blocks are then hauled in so that they are in contact with the hull. These side blocks are hauled in using chains and pulleys operated from the dock's wing walls. It is critical that this is done while the ship is coming out of the water prior to the ship reaching the draft of unstability. If not, the ship could roll off the keel blocks and overturn in the dry dock.

Fig. 3. USS *Los Alamos* (AFDB 7), a U.S. Navy floating sectional dry dock, at Holy Loch, Scotland (where it was located for many years), with a submarine docked in it. (*U.S. Navy*)

Fig. 4. British Royal Navy ship HMS *Halifax* in a graving dock in 1889 at Halifax, Nova Scotia, Canada. (*Courtesy of Irving Shipbuilding Inc.*)

An advantage of a floating dry dock is that it can be moved from shipyard to shipyard or towed across the sea to remote locations where dry docks are needed. The first forward-positioned floating dry dock of the U.S. Navy was the *Dewey*, which was towed from Chesapeake Bay on the United States east coast to Subic Bay, Philippines, in 1906 over a period of $6^1/_2$ months, by way of the Atlantic, the Mediterranean, the Suez Canal, the Indian Ocean, and the South China Sea.

A floating dry dock does need a place in the harbor to be submerged. This "hole" is usually an area dredged so that the floating dry dock can fully ballast down and the ship to be docked can enter the dry dock and be positioned over the keel blocks.

Graving dock. This is a dry dock dug out of the ground, usually with a thick concrete base supported by pilings and surrounded on three sides by earth held back by concrete walls anchored back into the surrounding soil. Sometimes graving dock walls are constructed of steel sheet piling structures. The seaward entrance usually has a gate which can be raised or lowered or a floatable vessel called a caisson which is ballasted down at the entrance into a sill to seal off the dock from the adjacent waterway. Large electrically driven pumps then pump out the graving dock. The dock is flooded via flooding tunnels by opening large gate-type valves.

When the dock is flooded and the gate is lowered or the caisson is removed, a ship is brought into the graving dock using boats and docking lines. The ship is precisely positioned over the blocking systems which is on the floor of the dock, and the gate is then raised or the caisson is ballasted down into the sill. The graving dock's dewatering pumps are then started and slowly the water in the dock drops. The ship lands on the blocking system, and the water continues to be pumped out until the dock floor is dry.

In **Fig. 4**, the HMS *Halifax* is seen on the blocks in a partially flooded graving dock in 1889, supported

Fig. 5. USS *Missouri* (BB 63) in Dry Dock #1, a graving dock at Long Beach Naval Shipyard in California (since closed), in 1984. (*U.S. Navy*)

Fig. 6. Graving dock being flooded at Pearl Harbor Naval Shipyard. (*U.S. Navy*).

on either side by what are called wale shores. These shores, which look like telephone poles on either side of the ship, are wedged into place to keep the ship from overturning in the dock. They are used instead of or in addition to side blocks.

Figure 5 shows the USS *Missouri* (BB 63) in a graving dock in 1984. The dry dock was built to handle aircraft carriers, but a battleship does a nice job of filling it. One advantage of a graving dock, seen here, is that the main deck of the ship is almost even with the ground level of the shipyard, making it very easy for shipyard workers to get on and off the ship. Also, the dockside portal cranes have easy reach of the entire ship.

Figure 6 shows a graving dock being flooded and a frigate's dry dock blocking system being covered by the water entering the dock. Soon the water will be high enough in the dock to float the ship off the blocks.

Graving dock dry-dockings are also inherently safer operations as no pumping plan is required. The structural limit of the graving dock still must be considered, however, when determining if the floor capacity of the dock can handle the weight of the ship and blocking system.

Vertical lift dock. This type of dry dock is somewhat similar to an elevator. The platform, supported by winch cables, is lowered into the water. A ship is floated in over the platform; then the platform is lifted using electrically powered winches all synchronized at the same speed. The ship lands on the blocking system and the platform is raised up out of the water.

The blocking system is on a cradle supported by wheels. The platform has a track system whereby the ship can be towed off the platform and moved to a work area in the shipyard (**Fig. 7**). This frees up the platform for another docking evolution.

Alternating-current synchronous induction motors are used to power the cable winches that raise and lower the ship platform. This enables the winches to all operate at the same lift speed regardless of load. The strain or load on each cable is monitored. The load on the cables is actually a measurement of the combined load distribution of the ship, the blocking and carriage system, and the vertical lift platform. Off-center weights can also be detected if a level platform shows different loads on the cables on either side of the lift.

Fig. 7. USS *Elliot* (DD 967) on a mechanical or vertical lift at the Todd Pacific Shipyard at San Pedro, California, around 1984. When the shipyard closed in the early 1990s, the vertical lift system was disassembled and moved to Malaysia, where it is now operating (as of 2005). (*Rolls-Royce Naval Marine Inc.*)

The advantage of this type of dock is that the ships, once docked, can be moved to another location, so that the vertical lift is freed up to conduct other dockings. This arrangement greatly increases the capacity of a shipyard. Sand blasting operations do require barriers to be erected to reduce the spread of the grit, and access to the ship by shipyard workers is via stairways or elevators, making it more difficult.

Hull preservation and maintenance. One of the most important maintenance actions while a ship is in dry dock is painting the hull. Usually the ship's hull is blasted with abrasives down to white metal and covered with one or more primer coats, a layer of anticorrosive paint, and two coats of antifouling paint. It is critical throughout this process that the temperature and humidity be monitored properly and that the required drying times are followed to prevent premature and catastrophic paint coating system failures later on.

Other critical work typically includes propeller replacement, pulling shafts and replacing journal bearing surfaces, refurbishing sea valves, cleaning the sea chests, and replacing transducers on the bottom. Another critical maintenance action in dry dock is the blasting and painting of all tanks that touch the underwater hull. It is very difficult to get a coating system to properly adhere to the hull plating when cool water is on the outside because the sweating of the hull inside surface prevents proper bonding and drying of the coating system. Richard D. Hepburn; Lemuel C. Robertson

Bibliography. R. D. Hepburn, *History of American Naval Dry Docks: A Key Ingredient to a Maritime Power*, Noesis, Inc., 2003; B. K. Mazurkiewicz, *Design and Construction of Dry Docks*, 1980.

Dry ice

A solid form of carbon dioxide, CO_2, which finds its largest application as a cooling agent in the transportation of perishables. It is nontoxic and noncorrosive and sublimes directly from a solid to a gas, leaving no residue. At atmospheric pressure it sublimes at −109.6°F (−78.7°C), absorbing its latent heat of 246.4 Btu/lb (573.1 kilojoules/kg). Including sensible heat absorption, the cooling effect per pound (kilogram) of dry ice is approximately 270 Btu (628.0 kJ) at storage temperatures above 15°F (−9°C) and 250 Btu (581.5 kJ) at lower temperatures. Slabs of dry ice can easily be cut and used in shipping containers for frozen foods, in refrigerated trucks, and as a supplemental cooling agent in refrigerator cars.

The manufacture of carbon dioxide gas is a chemical process. The gas is liquefied by compressing it to 900–1000 lb/in.^{2} (6.2–6.9 megapascals) gage in three stages of reciprocating compressors and then condensing it in water-cooled condensers. The liquid is expanded to atmospheric pressure where its temperature is below the triple point (−69.9°F or −56.6°C). In the expansion there is a flash separation (at −109.6°F or −78.7°C) of some solid carbon dioxide "snow" which is very porous. The snow is removed from the expansion chamber and mechanically compressed into standard 50-lb (23-kg) blocks which measure 10 by 10 by 10 in. (25 by 25 by 25 cm). *See* CARBON DIOXIDE. Carl F. Kayan

Drying

An operation in which a liquid, usually water, is removed from a wet solid in equipment termed dryers. The use of heat to remove liquids distinguishes drying from mechanical dewatering methods such as centrifugation, decantation or sedimentation, and filtration, in which no change in phase from liquid to vapor is experienced. Drying is preferred to the term dehydration, which is sometimes used in connection with the drying of foods. Dehydration usually implies removal of water accompanied by a chemical change. Drying is a widespread operation in the chemical process industries. It is used for chemicals of all types, pharmaceuticals, biological materials, foods, detergents, wood, minerals, and industrial wastes. Drying processes may evaporate liquids at rates varying from only a few ounces per hour to 10 tons per hour in a single dryer. Drying temperatures may be as high as 1400°F (760°C), or as low −40°F (−40°C) in freeze drying. Dryers range in size from small cabinets to spray dryers with steel towers 100 ft (30 m) high and 30 ft (9 m) in diameter. The materials dried may be in the form of thin solutions, suspensions, slurries, pastes, granular materials, bulk objects, fibers, or sheets. Drying may be accomplished by convective heat transfer, by conduction from heated surfaces, by radiation, and by dielectric heating. In general, the removal of moisture from liquids (that is, the drying of liquids) and the drying of gases are classified as distillation processes and adsorption processes, respectively, and they are performed in special equipment usually termed distillation columns (for liquids) and adsorbers (for gases and liquids). Gases also may be dried by compression. *See* ADSORPTION; DISTILLATION.

Drying of solids. In the drying of solids, the desirable end product is in solid form. Thus, even though the solid is initially in solution, the problem of producing this solid in dry form is classed under this heading. Final moisture contents of dry solids are usually less than 10%, and in many instances, less than 1%.

The mechanism of the drying of solids is reasonably simple in concept. When drying is done with heated gases, in the most general case, a wet solid begins to dry as though the water were present alone without any solid, and hence evaporation proceeds as it would from a so-called free water surface, that is, as water standing in an open pan. The period or stage of drying during this initial phase, therefore, is commonly referred to as the constant-rate period because evaporation occurs at a constant rate and is independent of the solid present. The presence of any dissolved salts will cause the evaporation rate to be less than that of pure water. Nevertheless, this

lower rate can still be constant during the first stages of drying.

A fundamental theory of drying depends on a knowledge of the forces governing the flow of liquids inside solids. Attempts have been made to develop a general theory of drying on the basis that liquids move inside solids by a diffusional process. However, this is not true in all cases. In fact, only in a limited number of types of solids does true diffusion of liquids occur. In most cases, the internal flow mechanism results from a combination of forces which may include capillarity, internal pressure gradients caused by shrinkage, a vapor-liquid flow sequence caused by temperature gradients, diffusion, and osmosis. Because of the complexities of the internal flow mechanism, it has not been possible to evolve a generalized theory of drying applicable to all materials. Only in the drying of certain bulk objects such as wood, ceramics, and soap has a significant understanding of the internal mechanism been gained which permits control of product quality.

Most investigations of drying have been made from the so-called external viewpoint, wherein the effects of the external drying medium such as air velocity, humidity, temperature, and wet material shape and subdivision are studied with respect to their influence on the drying rate. The results of such investigations are usually presented as drying rate curves, and the natures of these curves are used to interpret the drying mechanism. **Figure 1** shows a series of typical drying-rate curves.

The constant-rate period of drying when heat is supplied by convection is susceptible to theoretical and analytical treatment because it is essentially independent of the solid material. When drying is accomplished by heat transfer from hot gases, which also remove the evolved vapors, the constant rate may be expressed in terms of heat-transfer rates or mass-transfer rates.

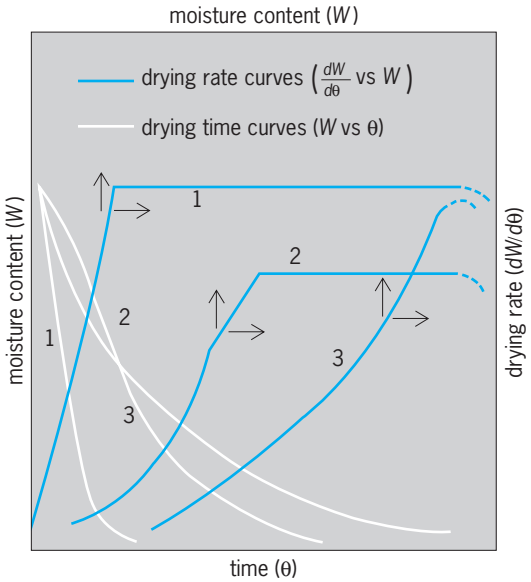

Fig. 1. Drying-time and drying-rate curves illustrating the general problem of drying: (1) Curves typical of a layer of thin material with most of the drying in the constant rate. (2) A more general case in which two stages in the falling-rate period occur. Typical of granular materials. (3) A case in which no constant rate occurs. Typical of homogeneous and colloidal materials such as soap, gelatin, and viscous solutions.

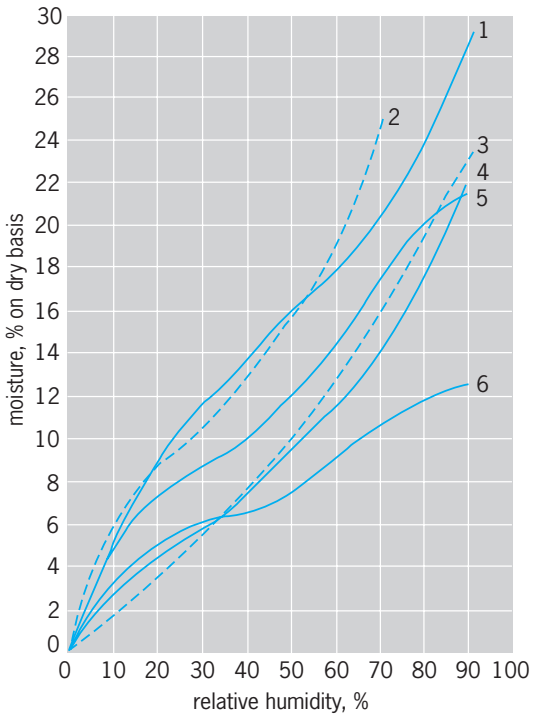

Fig. 2. Equilibrium moisture content of miscellaneous organic materials at 70°F (21°C): (1) leather; (2) tobacco; (3) soap; (4) wood; (5) catgut; (6) glue.

A constant rate of evaporation at the surface of the solid maintains the surface at a constant temperature, which, in the absence of other heat effects, is very nearly the wet-bulb temperature of the air. This temperature may range from 70 to 130°F (21 to 54°C) for convection drying, depending on the temperature and humidity of the air and on radiation. This so-called wet-bulb cooling effect is one reason why heat-sensitive solids can be dried in air at temperatures well above the decomposition temperature of the solid.

The magnitude of the constant rate can vary widely, depending on the degree of subdivision of the material, that is, the manner in which the material is exposed to the drying air. Thus, the rate of drying in spray dryers can be several hundred-thousand-fold greater than the rates in tray dryers.

A number of empirical expressions based on experimental studies have been developed for estimating the constant rate for different physical configurations of the wet material.

When materials are dried in contact with hot surfaces, termed indirect drying, the air humidity and air velocity may no longer be significant factors controlling the rate. The "goodness" of the contact between the wet material and the heated surfaces, plus the surface temperature, will be controlling. This may involve agitation of the wet material in some cases.

The falling-rate period is not as amenable to treatment as the constant-rate period because the falling

rate depends largely on the internal structure of the solid and the mechanism of moisture flow therein. In falling-rate processes, the rate of drying decreases gradually until the moisture content of the material approximates the equilibrium value. The equilibrium moisture content of a material is that moisture content to which a given material can be dried under specific conditions of air temperature and humidity. A typical equilibrium moisture curve is shown in **Fig. 2**. A unique characteristic of hygroscopic materials is that they hold or retain water at a vapor pressure less than water at the same temperature. Moisture so retained is termed bound moisture. Materials in which water exerts its normal vapor pressure at all moisture contents are termed nonhygroscopic, and in general, are easier to dry.

Classification of dryers. Drying equipment for solids may be conveniently grouped into three classes on the basis of the method of transferring heat for evaporation. The first class is termed direct dryers; the second class, indirect dryers; and the third class, radiant heat dryers. In the chart in **Fig. 3**, each class is subdivided into batch and continuous. Batch dryers are restricted to low capacities and long drying times. Most industrial drying operations are performed in continuous dryers. The large numbers of different types of dryers reflect the efforts to handle the larger numbers of wet materials in ways which result in the most efficient contacting with the drying medium. Thus, filter cakes, pastes, and similar materials, when preformed in small pieces, can be dried many times faster in continuous through-circulation dryers than in batch tray dryers. Similarly, materials which are sprayed to form small drops, as in spray drying, dry much faster than in through-circulation drying.

Direct dryers. The general operating characteristics of direct dryers are the following. (1) Drying is accomplished by convection heat transfer between the wet solid and a hot gas, the latter removing the vaporized liquid as well as supplying the heat needed for evaporation. (2) The heating medium may be steam-heated air, gases of combustion, a heated inert atmosphere such as nitrogen, or a superheated vapor such as steam. (3) Drying temperatures may range from prevailing atmospheric temperatures to 1400°F (760°C). (4) At drying temperatures below the boiling point of the liquid, increasing amounts of vapor of this liquid in the drying gas will decrease the rate of drying and increase the final liquid content of the solid. (5) When the drying temperatures are above the boiling point throughout the process, an increase in the vapor content of the gas or air, in general, will have no retarding effect on the drying rate, and no effect on the final moisture content. (6) For low-temperature drying, dehumidification of the drying gas is often required when high atmospheric humidities prevail. (7) The efficiency of direct dryers will increase with an increase in the inlet temperature of the drying gas at a fixed exhaust temperature.

Indirect dryers. The general operating characteristics of indirect dryers are as follows. (1) Drying by the transfer of heat by conduction and some radiation to the wet material; conduction usually occurs through a metallic retaining wall. The source of heat is generally condensing steam, but may also be hot water, gases of combustion, molten heat-transfer salts, hot oil, or electric heat. (2) The drying temperature of the surface of contact may range from below freezing to 1000°F (537°C). (3) Indirect dryers are especially suited to drying under reduced pressures and with inert atmospheres, and are therefore well adapted to the recovery of solvents. (4) Indirect dryers using condensing steam usually have a high efficiency because heat is supplied according to the demand, but, as in all dryers, the efficiency falls off markedly when very low final moisture contents are required. (5) Indirect dryers can handle dusty materials more readily than direct dryers. (6) The operation of indirect dryers is frequently characterized by some method of agitation to improve the contact between the hot metal surface and wet material. The nature of this contact determines the drying rate of indirect dryers. Therefore, heavy, granular materials generally give higher heat-transfer coefficients of contact than fluffy, bulky solids.

Radiant-energy dryers. These operate by the transfer of heat from a radiant source to the wet material being dried. The temperature of the radiant source may range from hot water or steam temperatures, 200–350°F (93–176°C), to the temperatures of incandescent surfaces, 1500–2500°F (815–1371°C). The generating medium may be steam, hot liquids, gas flames, or electricity, depending on the temperature desired and the equipment design.

Special types. Dielectric-heat dryers do not fall in any of the above classes inasmuch as their operation depends on the generation of heat by high-frequency fields inside the material being dried so that heat will actually flow out of the interior of the solid. These dryers are used to dry large bulky objects which have a long internal path for moisture flow.

Direct batch dryers. In operation, heated air circulates over the wet material being dried. The wet solid is supported according to its physical form. Lumber, ceramics, and similar massive objects are stacked in piles or on racks; textile skeins, painted objects, and hides are suspended on hangers; and granular materials, pastes, slurries, and liquids are placed in trays, which may be supported on stationary or movable racks. Good performance by this type of dryer depends on uniform, equal air velocities across all the wet material.

In batch–through-circulation dryers, heated air is blown through the wet material on screen-bottom trays instead of across. The material to be dried must be permeable to air flow.

Dryers of this type, but of binlike design, are used extensively in the explosives industry to dry gunpowder, and in food processing to dry and condition certain foodstuffs, such as grains and corn.

Direct continuous dryers. Tunnel and continuous tray dryers usually consist of long, enclosed housing or tunnels through which wet material is moved on trucks. Hot air is blown through the trucks. Air

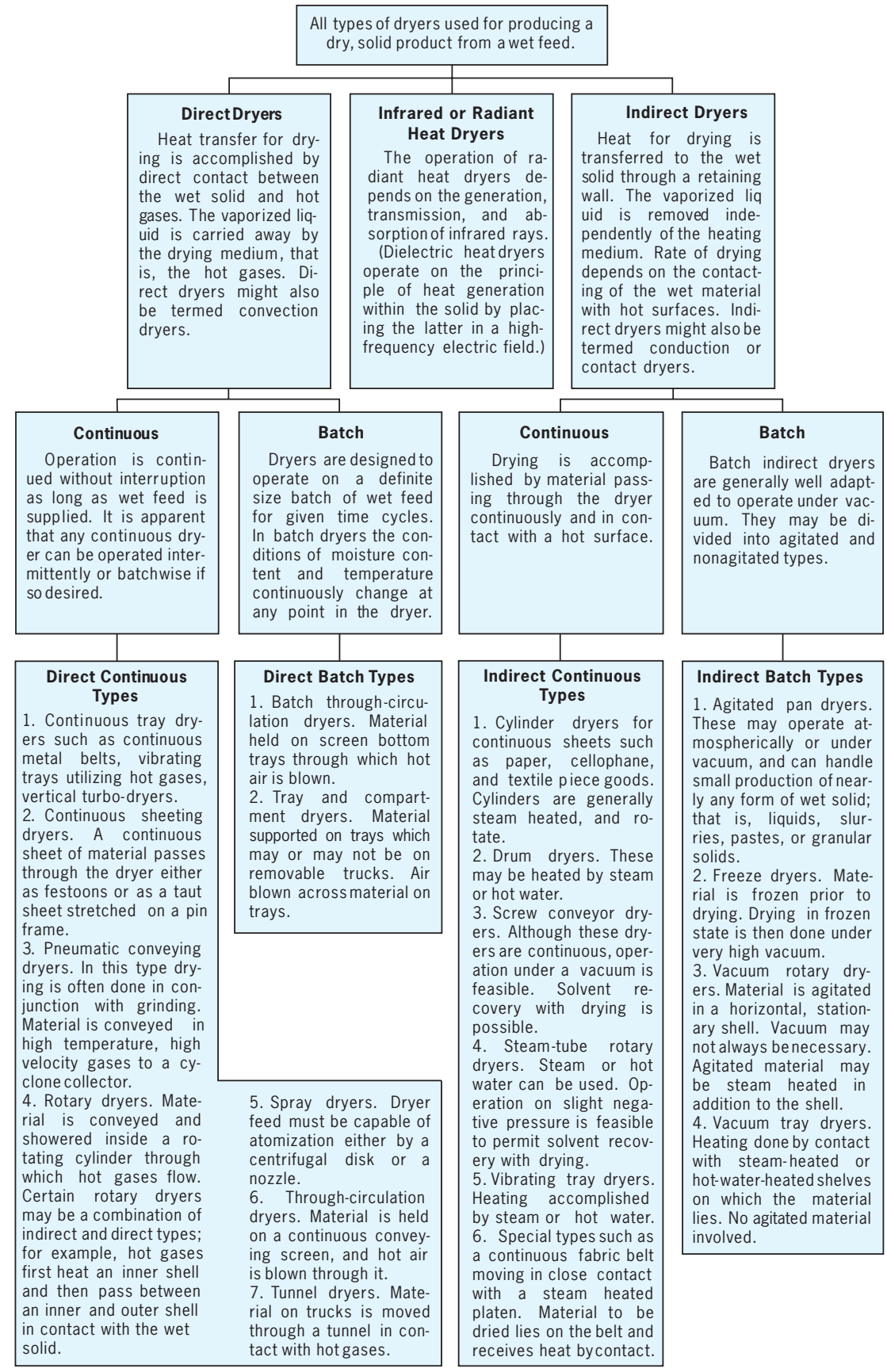

Fig. 3. Classification of dryers, based on methods of heat transfer.

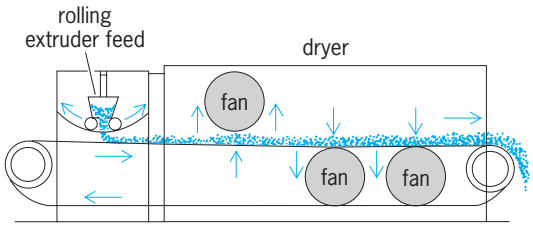

Fig. 4. Diagram of path of travel of permeable bed of wet material through a three-unit through-circulation dryer. (*Proctor and Schwartz, Inc.*)

flow may be parallel, counter flow, or at right angles (cross flow) to the movement of the trucks. The trucks may move continuously or semicontinuously through the tunnel. Tunnel dryers may be operated adiabatically, that is, without the addition of heat in the tunnel, or the air may be reheated periodically during its passage through the tunnel.

Wet granular materials are held on the trays of trucks; foodstuffs, rayon cakes, pottery, and large ceramic objects are held on racks; textile skeins are draped over rods; and hides are pasted on glass plates or hung on frames.

In continuous through-circulation dryers, heated air is blown through a permeable bed of wet material as it passes continuously through the dryer. Drying rates are much higher than in the usual tray or tunnel dryers because of the large surface area exposed per unit weight of material and because the smaller particles offer less resistance to internal moisture flow.

The operation of this type of dryer depends on whether or not the wet material is in a state of subdivision suitable for through circulation of the hot air. Some materials are already in such a permeable state. Many materials require special preliminary treatment, termed preforming, to form them into permeable beds. Preforming may include the processes of scoring on rotary filters, granulating, extruding, briqueting, flaking, and predrying on finned drums.

One type of through-circulation dryer consists of a horizontal conveying screen, which moves through a tunnellike housing. A permeable bed of the wet material is supported and conveyed on the screen, and hot air or gas is blown vertically up or down through the bed (**Fig. 4**). Through-circulation drying may also be performed in rotary-type dryers, which convey the material by a tumbling action imparted by the rotation of the dryer shell.

Conveying-screen through-circulation dryers are widely used in chemical plants to dry fibrous, flaky, and granular materials, such as cotton linters, rayon staple, cellulose acetate, silica gel, sawdust, and minerals, materials that are well suited to through circulation of air without prior treatment. They are also used extensively to dry materials, such as starch, pigments, calcium carbonate, insecticides, dyes, and intermediates, which must be preformed by one of the methods noted above.

Direct rotary dryers are used extensively in the chemical industry. A rotary dryer consists of a cylinder slightly inclined to the horizontal and rotated on suitable bearings. The rotary action of the cylinder serves to convey wet material from one end to the other while passing hot gases axially through the dryer shell. The contact of the solids and gases is further improved by means of flights arranged within the dryer shell so as to shower the wet material through the hot gas stream. A rotary dryer without such lifting flights is usually called a rotary kiln.

Rotary dryers may operate with air flow either parallel or countercurrent to the flow of the wet solid (**Fig. 5**).

Pneumatic conveying dryers, sometimes termed flash or dispersion dryers, operate on the basis of simultaneously conveying and drying a wet solid in a high-velocity stream of hot gas. Temperatures up to 1400°F (760°C) are used in these dryers. The short contact times involved permit usage gas temperatures above the decomposition temperature of the material. The gas stream acts as both the conveying and heating medium. Gas velocities on the order of 75 ft/s (23 m/s) are used. Frequently, the wet material is in such a form that some disintegrating action is required before it can be conveyed. A schematic diagram of this type of dryer is shown in **Fig. 6**. It is applicable to granular, free-flowing materials, such as coal, whey, and sodium chloride, and to sludges, filter-press cakes, and similar nongranular solids which require disintegration for proper dispersion. It is common practice to recycle dry product

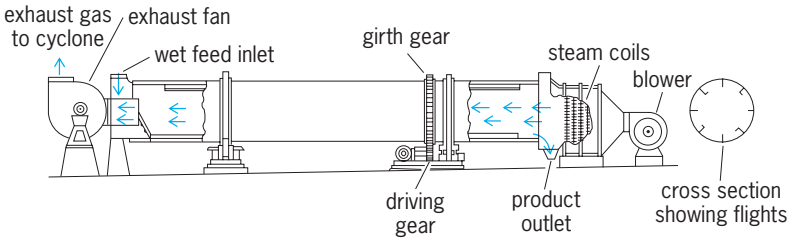

Fig. 5. Single-shell direct rotary using steam-heated air and balanced pressure by means of a blower and exhauster. (*Hardinge Co.*)

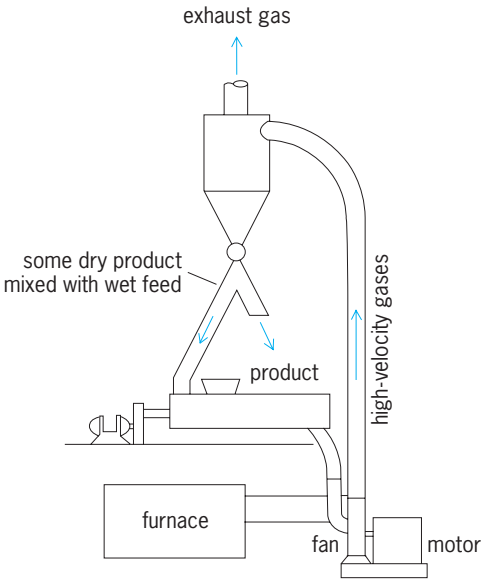

Fig. 6. Dispersion dryer.

into the wet feed in order to facilitate dispersion and handling.

Spray dryers operate on the principle of creating a highly dispersed liquid state in a high-temperature (up to 1400°F or 760°C) gas zone. The heart of a spray-drying process is the creation of small liquid droplets by spraying. This may be accomplished by means of (1) high-pressure nozzles, (2) pneumatic nozzles, and (3) high-speed rotating disks. Almost any pumpable liquid, from a thin clear liquid to a pasty sludge, can be atomized sufficiently for spray drying. Liquids above a viscosity of 1500 centipoises, however, are very difficult to atomize and spray dry. Generally, fine atomization will not produce a large percentage of droplets less than 5 micrometers in diameter. The particle size under conditions of so-called coarse atomization will be on the order of 200–600 μm. Because of the high surface-volume ratio of small drops, the actual drying time in spray dryers may be considerably less than 1 s for high-temperature operation.

Spray dryers are made in a multitude of designs. The drying gases may flow cocurrent with or countercurrent to the spray, and the spray may be directed vertically up or down, or horizontally. Various types of spray dryers are shown in **Fig. 7**. They are used extensively in the chemical, pharmaceutical, and food industries.

Fluidized-bed dryers operate by having heated air pass upward at sufficient velocity through a column or layer of granular, wet material to cause it to fluidize and become mixed by turbulent action. Wet feed may be introduced at the bottom of the column, and dry product removed from the top, as is shown in **Fig. 8**.

In direct continuous-sheeting dryers, heated air is circulated over or through a continuously moving wet sheet, which is supported by a variety of methods.

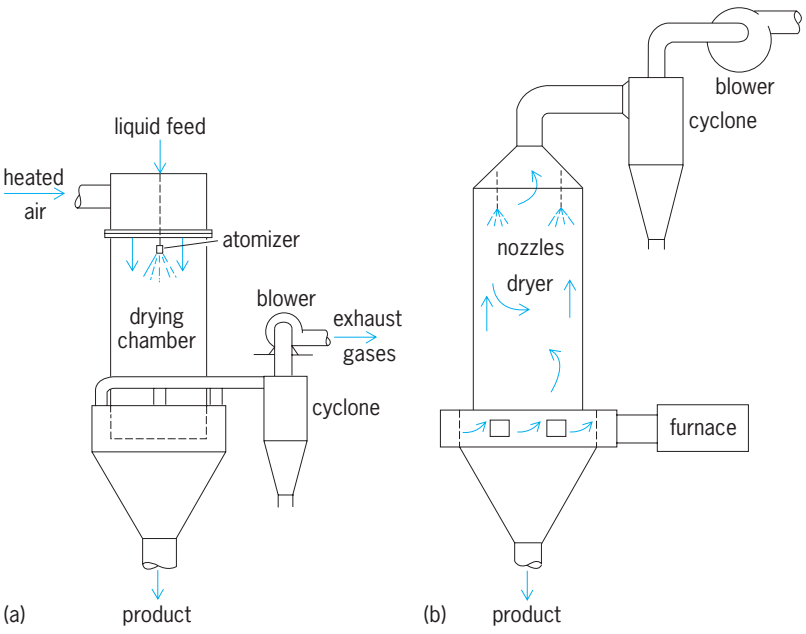

Fig. 7. Spray dryers. (*a*) Cocurrent. (*b*) Countercurrent.

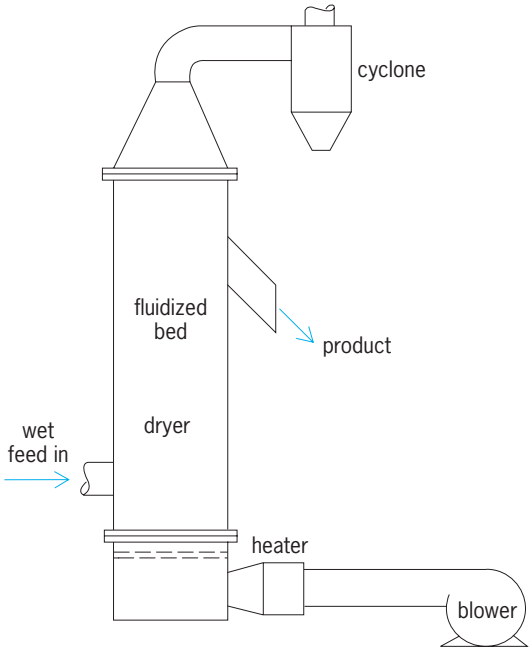

Fig. 8. Fluidized-bed dryer.

Direct continuous-sheeting dryers occur in a variety of types. The festoon or loop dryer permits drying of continuous sheet material in a relaxed state by festooning or looping it over rolls, which in turn are conveyed through the dryer. Tenter dryers are used to dry a continuous sheet of material under tension. Heated air is usually blown perpendicularly from slots or nozzles against both sides of the sheet.

Direct continuous-sheeting dryers are used to dry sheet materials that are sufficiently strong mechanically in the wet and dry states to support their own weight or withstand tension as required. Their widest use is in the textile field, in the manufacture of coated and impregnated fabrics, and in the preparation of films and coated papers.

Indirect batch dryers. Vacuum shelf dryers are indirect batch dryers which generally consist of a vacuum-tight cubical or cylindrical chamber of cast-iron or steel plate, heated supporting shelves inside the chamber, a vacuum source, and a condenser. In operation, heat transfer takes place by conduction through metal surfaces and interfaces to the wet material held on the shelves.

Vacuum shelf dryers are used extensively for drying pharmaceuticals, temperature-sensitive or easily oxidizable materials, and small batches of high-cost products where any product loss must be avoided. This type is particularly useful for recovery of valuable solvents or vapors.

Vacuum freeze dryers are used principally for drying materials that would be destroyed by the loss of volatile ingredients or by drying temperatures above the freezing point. The material is dried in the frozen state, so that a process of sublimation is involved, that is, ice sublimes directly to water vapor. Because the material dries in a rigid frozen condition, shrinkage is minimized, and the resulting structure of the dry solid is usually porous and readily soluble. This led

to the term lyophilization in the early development of freeze drying.

The equipment required for this method of drying consists of a heated drying chamber where sublimation occurs, a piping system for the transport of vapors, and a vapor-removal system for condensable and noncondensable gases. Removal of the condensable vapors may be accomplished by freezing on cold surfaces, absorption in liquid desiccants, or adsorption on solid desiccants.

One application of freeze drying (the Cryochem process) involves conduction heat transfer to the frozen solid held on a metallic surface. However, should the metal surface temperature rise above the freezing point of the solid, melting will occur. A second method of freeze drying utilizes heat transfer by radiation.

Agitated-pan dryers consist of a bowl- or pan-shaped receptacle in which wet material is stirred or agitated in contact with hot surfaces. The operation may be atmospheric or vacuum. Agitated-pan dryers are used to handle small batches of materials that can withstand agitation during drying. They are suitable for pastes and slurries, and for materials containing valuable solvents which must be recovered.

The vacuum rotary dryer is another type of batch indirect dryer with an agitator. It consists of a horizontal shell in which agitator blades attached to a horizontal rotating shaft revolve and agitate the material being dried (**Fig. 9**). Heat is supplied by condensing steam in a jacket surrounding the shell, by hot water, or by other heat-transfer fluids. Vacuum rotary dryers are used for large batches of materials that must be dried in the absence of air or where the recovery of solvents is required.

Indirect continuous dryers. Indirect rotary dryers are similar mechanically and in appearance to the direct rotary dryers discussed above. They differ primarily in that heat is transferred to the material through the metal shell or from steam tubes located around the dryer shell, rather than from hot gases as in the case of direct rotary dryers.

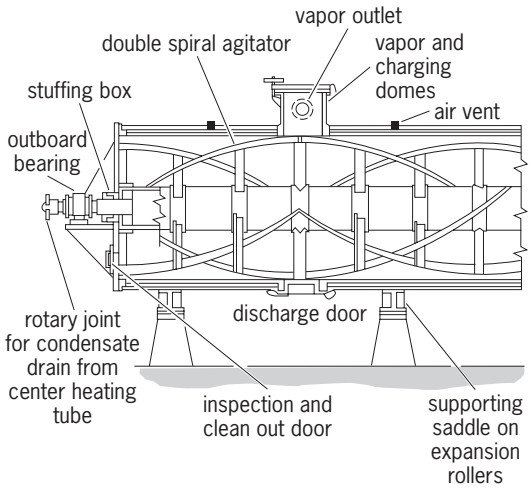

Fig. 9. Typical vacuum rotary dryer. (*Blaw-Knox Co.*)

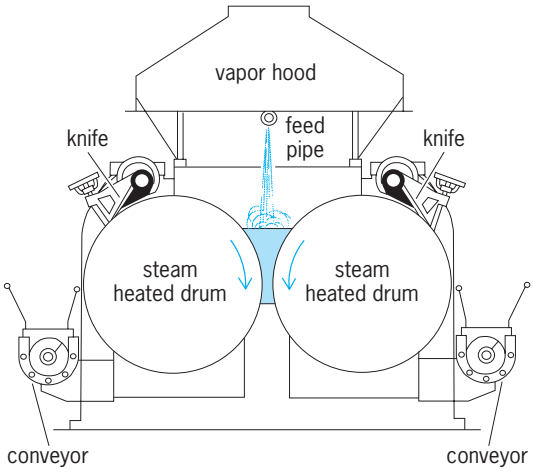

Fig. 10. Double-drum dryer with pipe feed. (*Buflovak Equipment Division, Blaw-Knox Co.*)

A screw-conveyor dryer is essentially a jacketed conveyor in which material is heated and dried as it is conveyed. In one type, the jacket may extend only to the top of the conveyor, which is left open to the atmosphere. This is termed a trough dryer. When the jacket encloses the conveyor completely, a slight negative pressure is required to sweep out the evaporated moisture.

Vibrating-conveyor dryers consist of a vibrating, heated, solid deck over which the wet solid moves. This deck can be heated by hot gases, steam, methanol vapors, or Dowtherm vapors, which pass through a jacket fastened to the deck upon which the material is conveyed. Direct gas flames can also be used as the source of heat. A hood equipped with an exhaust fan and placed over the deck removes the evaporated liquid. Infrared lamps may be mounted above the deck to increase drying rates.

Drum drying consists of applying a liquid material, solution, slurry, or paste to a revolving heated metal drum, which conducts heat to the wet film to evaporate the water during a partial revolution of the drum (**Fig. 10**). The dry material is scraped from the drum by a stationary knife. Drum dryers may be designated by type as atmospheric double-drum, atmospheric single-drum, atmospheric twin-drum, vacuum single-drum, and vacuum double-drum dryer.

In drum drying, the product is exposed to heat for only short periods of time. This has the advantage that, although the product may approach the temperature of the drum surface, there usually is no adverse effect from overheating.

Cylinder dryers, sometimes called can dryers or drying rolls, differ from drum dryers in that they are used for materials in a continuous-sheet form. Cylinder dryers may consist of one large cylindrical drum, such as the so-called Yankee dryer; more often they comprise a number of drums arranged so that a continuous sheet of material can pass over them in series. Typical of this arrangement are Fourdrinier paper-machine dryers, cellophane dryers, and slashers for textile piece goods and fibers.

The size of commercial cylinder dryers covers a wide range. The individual rolls may be 2–6 ft (0.6–1.8 m) in diameter and up to 20 ft (6 m) in width. In some cases, the width of the rolls decreases throughout the dryer in order to conform to the shrinkage of the sheet.

Drying of gases. The removal of 95–100% of the water vapor in air or other gases is frequently necessary. Gases having a dew point of −40°F (−40°C) are considered commercially dry. The more important reasons for the removal of water vapor from air are (1) comfort, as in air conditioning; (2) control of the humidity of manufacturing atmospheres; (3) protection of electrical equipment against corrosion, short circuits, and electrostatic discharges; (4) requirement of dry air for use in chemical processes where moisture present in air adversely affects the economy of the process; (5) prevention of water adsorption in pneumatic conveying; and (6) as a prerequisite to liquefaction.

Gases may be dried by the following processes: (1) absorption by use of spray chambers with such organic liquids as glycerin, or aqueous solutions of salts such as lithium chloride, and by use of packed columns with countercurrent flow of sulfuric acid, phosphoric acid, or organic liquids; (2) adsorption by use of solid adsorbents such as activated alumina, silica gel, or molecular sieves; (3) compression to a partial pressure of water vapor greater than the saturation pressure to effect condensation of liquid water; (4) cooling below dew point of the gas with surface condensers or coldwater sprays; and (5) compression and cooling, in which liquid desiccants are used in continuous processes in spray chambers and packed towers—solid desiccants are generally used in an intermittent operation that requires periodic interruption for regeneration of the spent desiccant.

Desiccants are classified as solid adsorbents, which remove water vapor by the phenomena of surface adsorption and capillary condensation (silica gel and activated alumina); solid absorbents, which remove water vapor by chemical reaction (fused anhydrous calcium sulfate, lime, and magnesium perchlorate); deliquescent absorbents, which remove water vapor by chemical reaction and dissolution (calcium chloride and potassium hydroxide); or liquid absorbents, which remove water vapor by absorption (sulfuric acid, lithium chloride solutions, and ethylene glycol).

The mechanical methods of drying gases, compression and cooling and refrigeration, are used in large-scale operations, and generally are more expensive methods than those using desiccants. Such mechanical methods are used when compression or cooling of the gas is required.

Liquid desiccants (concentrated acids and organic liquids) are generally liquid at all stages of a drying process. Soluble desiccants (calcium chloride and sodium hydroxide) include those solids which are deliquescent in the presence of high concentrations of water vapor.

Deliquescent salts and hydrates are generally used as concentrated solutions because of the practical difficulties in handling, replacing, and regenerating the wet corrosive solids. The degree of drying possible with solutions is much less than with corresponding solids; but, where only moderately low humidities are required and large volumes of air are dried, solutions are satisfactory. *See* DESICCANT; EVAPORATION; FILTRATION; HEAT TRANSFER; HUMIDIFICATION; UNIT OPERATIONS; VAPOR PRESSURE.

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Drying oil

A relatively highly unsaturated oil which, on exposure to air, is oxidized and polymerized to form a hard, dry film. Vegetable and fish oils are classified as nondrying, semidrying, or drying, according to their ease of autoxidation and polymerization. Due to their film-forming properties, drying oils are used in paints, inks, and varnishes. The reactivity of drying oils with oxygen results from the presence of diallylic groups (that is, two double bonds separated by methylene groups, —CH═CHCH$_2$CH═CH—) or conjugated double bonds (two carbon-carbon double bonds separated by a single bond). The number of diallylic groups *fn* per molecule is correlated with the oils' drying nature. Oils with *fn* greater than 2.2 are drying oils; those below 2.2 are semidrying to nondrying. The **table** shows the percentage of saturated and unsaturated fatty acids and the *fn* for some drying oils. These oils vary considerably by source and harvest time (see "Sunflower" in the table). *See* ALKENE; CONJUGATION AND HYPERCONJUGATION; FAT AND OIL; INK; VARNISH.

Studies of the reactions of ethyl linoleate with oxygen and driers (materials added to facilitate oxidation of oils) showed that cross-linking reactions mostly formed ether and peroxy cross-links with only about 5% of the cross-links being new C-C bonds. Polymerization of drying oils is accelerated at the surface with cobalt driers which accelerate the decomposition of hydroperoxides formed by oxidation. Zirconium is used as a through-film drier. The rearrangement and cleavage of hydroperoxides to aldehydes and ketones, during drying, causes much of the characteristic odor of oil and alkyd paints. *See* DRIER (PAINT); PAINT.

Drying oils are seldom used alone and primarily are raw materials for other binders such as oil-resin varnishes, heat-bodied oils (heated in an inert atmosphere at 300–320°C or 572–608°F), blown oils (heated in air at 150°C or 302°F), alkyd (polyester) resins, epoxy esters, and uralkyds (urethanes). These materials can be considered synthetic drying oils since *fn* has been raised above 2.2 by a chemical reaction. Alkyds are polyester resins formed by reacting vegetable oil or fatty acid with an aromatic diacid such as phthalic acid and a polyol such as

Percent fatty acid composition of selected oils

Name	Saturated[a]	Oleate, etc.	Linoleate	Linolenate	Other	*fn*
Castor	3	7	3		87[b]	0
Dehydrated castor	5	10	85			Conjugated
Coconut	91	7	2			0
Cottonseed	25	40	35			1
Fish oil (cod)					38[c]	6.9
Linseed	10	18	17	55		3.8
Oiticica	10	6	10		74[d]	4.8
Perilla	7	14	16	63		4.3
Safflower	11	13	75	1		2.3
Soybean	14	26	52	8		2.1
Sunflower (Minn.)[g]	13	26	61	Trace		1.8
Sunflower (Texas)	11	51	38	Trace		1.2
Tall oil fatty acids	8	46	41	3	2[e]	1.4
Tung	5	7	3		85[f]	Conjugated

[a] Saturated fatty acids are mainly mixtures of stearic (C_{18}) and palmitic (C_{16}) acids; coconut oil also contains C_8, C_{10}, C_{12}, and C_{14} saturated fatty acids.
[b] Ricinoleic acid.
[c] Clupanodonic (C22:6).
[d] Licanic.
[e] Rosin.
[f] α-Eleostearic acid (conjugated).

glycerin. Polyunsaturated acids dimerize by heat treatment with an acid catalyst. Dimer acids are predominantly C_{36} dicarboxylic acids. They are used in making polyesters and polyamides. *See* POLYAMIDE RESINS; POLYESTER RESINS. John L. Massingill, Jr.

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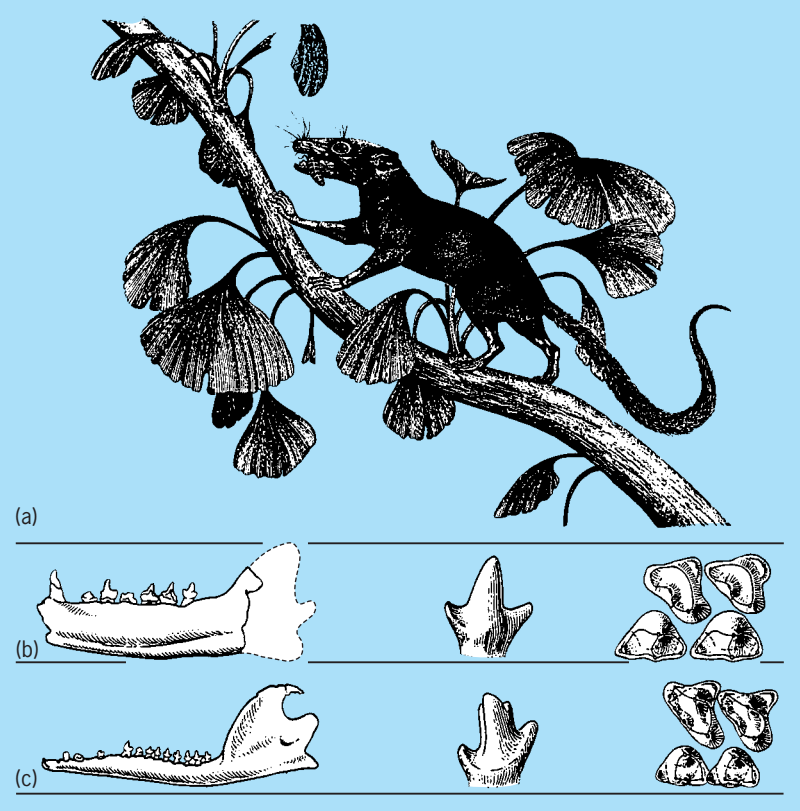
Dryolestida

The Dryolestida (formerly known as the Pantotheria or Eupantotheria) are an extinct order of mammals related to the living marsupials and eutherian (placental) mammals. Along with the closely related symmetrodonts, they were the most diverse group of mammals during the Jurassic. They are often found in some of the same quarries that yield the giant sauropod dinosaurs. Most were about the size and shape of a shrew or mouse, with teeth adapted for chopping up a diet of insects, and they apparently were adapted for living in trees like squirrels. *See* EUTHERIA; MARSUPIALIA; SYMMETRODONTA.

The earliest known dryolestoid is *Amphitherium*, first discovered in the Middle Jurassic Stonesfield Slate of England in 1838. This was an important specimen for historic reasons, proving that mammals once coexisted with the dinosaurs (at a time when it was thought that dinosaurs had died out before mammals appeared). By the Late Jurassic, dryolestoids were the most common mammals in both North America and Europe. They had seven genera of the family Dryolestidae in Europe and six in North America, as well as nine genera of the family Paurodontidae (mostly in North America). By the Early Cretaceous, dryolestids had been replaced in the Northern Hemisphere by the earliest relatives of marsupials and placentals, but they persisted until the Late Cretaceous of South America.

Of the 26 currently known genera of dryolestoids, all but one (*Henkelotherium*) are known only from remnants of teeth and jaws (see **illus.**). The teeth are very distinctive, with a number of unique specializations. The upper and lower molar teeth were made of tall triangles of three cusps, connected by crests between them. These molars were very compressed in a front-to-back direction, allowing as many as seven to nine molars in the cheek tooth row (marsupials typically have four molars; placentals have only three). The upper molar triangles (trigons) interlocked with the lower molar triangles (trigonids) like the teeth on a zipper, so that shear developed between the molars when the jaw closed. This type of dentition is similar to that found in living insectivorous mammals, and strongly suggests that dryolestoids were insectivorous as well. The upper molars have a large cusp on the exterior edge, which is called the stylocone, and another large cusp at the internal apex of the triangle, the paracone. In more advanced (tribosphenic) mammals such as marsupials and placentals, the paracone shifts to the outer edge and replaces the stylocone, and a new cusp, the protocone, develops in its place. The lower molar trigonids have a small ledge, the talonid, on the back end of the tooth, which occludes with the basin of the upper molar trigon. The paurodonts had a very similar dentition, but the cusps were lower and blunter. The teeth were less compressed in a front-to-back direction, and there were only four or five molars.

For over a century, dryolestoids were known only from dental remains. However, in 1991 a complete articulated skeleton of the paurodont *Henkelotherium* from the Late Jurassic beds of the Guimarota coal mine in Portugal was described. This skeleton was



Dryolestids. (*a*) Reconstruction of *Henkelotherium* (*after Krebs, 1991*). (*b*) Side view of jaw, side view of teeth, and crown view of upper and lower molars of the paurodonts *Paurodon* and *Pelicopsis*; (c) side views for the dryolestids *Laolestes* and *Melanodon* (*after Prothero, 1981*).

about the size of a small mouse, but was clearly adapted for living in trees, having long clawed fingers for gripping bark and a long, slender tail for balance. The hip and shoulder bones were remarkably modern and like those of most living mammals. Many of the reptilian bones of the shoulder girdle (still found in primitive mammals such as the egg-laying platypus, and some other Mesozoic mammals) were lost. However, the marsupial bones, which project forward from the pelvis in marsupials and help support the pouch, were retained. In addition, rudimentary vestiges of some of the reptilian jaw bones (such as the coronoids and splenials) were still present, although most of the lower jaw was still made of a single bone, the dentary, as in most other mammals. *See* DENTITION.

Dryolestoid diversity peaked in the Late Jurassic, and only one Early Cretaceous dryolestid (*Crusafontia* from Spain) was known. Then, in 1991, a peculiar form, *Donodon*, was described from the Early Cretaceous of Morocco. Between 1986 and 1992, a number of new dryolestoids from the Late Cretaceous Los Alamitos Formation of Argentina were described. At least four different families are represented, and they show a much greater variety of types of teeth than are known from the Jurassic. They include normal-looking dryolestids (*Leonardus* and *Groebertherium*) like those known from the Late Jurassic, as well as the much larger, blunt-cusped *Mesungulatum* (originally mistaken for an ancient hoofed placental mammal) and *Reigitherium*, with its flat-crowned teeth. These specimens show that dryolestoids lingered on in South America long after the northern continents were dominated with marsupials and placentals, and long after the dryolestid heyday in North America and Europe in the Late Jurassic and Early Cretaceous had passed.

For over a century, dryolestoids were placed in the order Pantotheria, a taxonomic wastebasket for Mesozoic mammals that were not members of other groups. In the older literature, one can still find statements that the "pantotheres" were ancestral to marsupials and placentals. Since 1981, the taxon Pantotheria has been abandoned, because it was a mixture of unrelated animals, and also because there was a tendency to use it uncritically as an ancestral group without detailed analysis of its members. Recent research has shown that the dryolestoids were primitive in many features but highly advanced in others. Their complex mosaic of features showed the transition from archaic mammals such as the platypus and some extinct Mesozoic mammals (which still retain reptilian characteristics) to the marsupials and placentals. In addition, their unique specializations (such as seven to nine highly unusual molars) show that they were not ancestral to marsupials or placentals, but an early side branch of mammalian evolution that lived before and then contemporaneously with the earliest marsupials and placentals. *See* THERIA.

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Duality (physics)

The state of having two natures, which is often applied in physics. The classic example is wave-particle duality. The elementary constituents of nature—electrons, quarks, photons, gravitons, and so on—behave in some respects like particles and in others like waves.

Duality is often used in a more precise sense. It indicates that two, seemingly different, theoretical descriptions of a physical system are actually mathematically equivalent. Such an occurrence is very useful. Various properties and phenomena are clearer in one or the other of the descriptions, and calculations that are difficult or impossible in one description may be simple in the other. In the case of wave-particle duality, the wave description corresponds to a theory of quantized fields, where the field variables are governed by an uncertainty principle. The particle description corresponds to a Feynman integral over all particle paths in spacetime. The quantized-field and path integral theories sound very different but

are mathematically equivalent, making identical predictions. *See* FEYNMAN INTEGRAL; NONRELATIVISTIC QUANTUM THEORY; QUANTUM FIELD THEORY; QUANTUM MECHANICS; UNCERTAINTY PRINCIPLE; WAVE MECHANICS.

Wave-particle duality is a very general principle, but more specific dualities arise in particular systems. These were once believed to occur primarily in rather simple or model systems, but they have been discovered in the very theories that are of interest in formulating a unified description of the particle interactions, gravity, and quantum mechanics. These new dualities have led to many surprising insights.

Weak-strong duality. The strengths of the interactions in any system are governed by one or more coupling constants. A typical coupling constant may be denoted by the symbol g. When g is small, the interactions are weak. It is then possible to make the approximation that in any process the interaction acts only once, with a small correction from two interactions, and so on: this is known as perturbation theory. However, when g is large, perturbation theory is inaccurate, and it can miss important effects. For example, such phenomena as quark confinement and superconductivity cannot be studied with perturbation theory. *See* PERTURBATION (QUANTUM MECHANICS); QUARKS; SUPERCONDUCTIVITY.

In some systems, there is weak-strong duality, meaning that when the coupling constant g of the original description is large that of the dual description, g', is small; for example $g' = 1/g$. When g is large, so the interactions in the original description are strong and the perturbation theory in this description is highly inaccurate, then perturbation theory in the dual description gives a very accurate description.

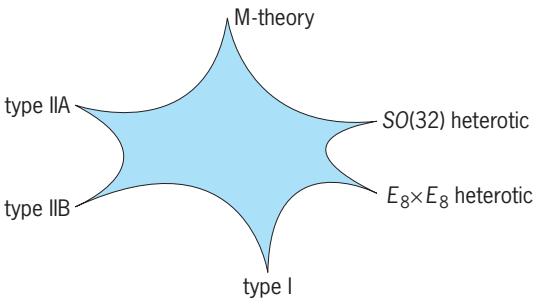
Electric-magnetic duality. A special form of weak-strong duality is electric-magnetic duality. Maxwell's equations, governing the behavior of the electromagnetic field, have a striking symmetry between the electric and magnetic fields, except that electric field lines can end on electric charges while magnetic field lines cannot end—there are no magnetic charges. The suggestion is that magnetic charges (also called magnetic monopoles) should exist, thus making the equations more symmetric. In fact, almost all theories that unify electromagnetism with the nuclear or gravitational interactions predict that such magnetic charges can exist. Generally the magnetic charges are superheavy, explaining why they have not been seen in particle accelerators. They might also have been produced in the big bang; to explain why they have not been seen was one of the motivations for inflationary cosmology. *See* BIG BANG THEORY; INFLATIONARY UNIVERSE COSMOLOGY; MAGNETIC MONOPOLES; MAXWELL'S EQUATIONS.

The electric and magnetic charges in these unified theories are rather different from one another. The electric charges are described as points, like the electron. The magnetic charges, however, arise as solitons—fuzzy lumps of field. Moreover, the electric charges are light and weakly coupled, while the magnetic charges are heavy and strongly coupled. To be precise, this is the situation when the coupling g is small. The Montonen-Olive conjecture is that when the coupling g becomes large, then in certain theories the picture reverses: the electric charges become fuzzy, heavy, and strongly coupled, while the magnetic charges become pointlike, light, and weakly coupled. Thus there is a dual description, with magnetically rather than electrically charged fields. The solitons of one description are the pointlike particles of the other. *See* SOLITON.

There is no proof of the Montonen-Olive conjecture, because adequate methods to analyze the large-g theory still do not exist. However, there is compelling circumstantial evidence that it is true. An important point is that the conjecture applies specifically to unified theories with supersymmetry. This is a symmetry between fermionic and bosonic particles, and is believed to be a necessary part of a fully unified theory. Supersymmetry has an interesting mathematical structure, which allows certain properties (some of the masses and interactions) of the large-g theory to be obtained exactly. This does not give a complete solution of the theory, but in combination with other physical input it gives a fairly detailed picture of the physics, and in particular gives strong evidence for Montonen-Olive duality. *See* SUPERSYMMETRY.

Duality in superstring theory. It is believed that a complete theory of all particles and interactions must be based on quantization of one-dimensional objects (loops) rather than points: this is superstring theory. In superstring theory there is again the problem that perturbation theory is the main tool, giving an incomplete description of the physics. The situation has greatly improved with the discovery that weak-strong duality is a general property of string theory. In fact, there are five known string theories, and all are dual to one another. A notable feature in string theory is that in addition to strings and solitons, duality requires certain other objects as well: D-branes, which are local disturbances to which strings become fixed.

The current understanding (*see* **illus.**) is that there is a single theory, with several coupling constants (which are actually spacetime fields). In different limits the physics is described by one of the weakly coupled string theories, or by something new—a



The five string theories and M-theory as limits of a single theory. The shaded region is parametrized by two coupling constants.

highly symmetric but only partially understood theory known as M-theory. The situation is analogous to the phase diagram for water (H_2O), where the different phases (solid, liquid, and gas) appear under different conditions, in analogy to the different string theories. The methods based on supersymmetry and duality have allowed an accurate mapping of the phase diagram of string theory, and have led to the discovery of new phases such as M-theory as well. *See* THERMODYNAMIC PROCESSES.

Remarkably, the same methods have also been used to solve some long-standing problems regarding the quantum mechanics of black holes. These solutions include a microscopic understanding of the analogy between black hole mechanics and thermodynamics, and progress toward the resolution of the information problem (an apparent conflict between quantum mechanics and general relativity). A further development is Maldacena duality, between black holes in string theory and ordinary quantum field theories with many fields. This has led to new understandings in field theory, string theory, and quantum gravity. *See* BLACK HOLE; QUANTUM GRAVITATION; SUPERSTRING THEORY. Joseph Polchinski

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Dubnium

A chemical element, symbol Db, atomic number 105. It was synthesized and identified unambiguously in March 1970 at the heavy-ion linear accelerator (HILAC) at the Lawrence Radiation Laboratory, Berkeley, University of California. The discovery team consisted of A. Ghiorso and colleagues. *See* PERIODIC TABLE.

The dubnium isotope, with a half-life of 1.6 s, decayed by emitting alpha particles with energies of 9.06 (55%), 9.10 (25%), and 9.14 (20%) MeV. It was shown to be of mass 260 by identifying lawrencium-256 as its daughter by two different methods.

Previous work on dubnium was reported in 1968 by G. N. Flerov and colleagues at Dubna Laboratories in Russia. They claimed to have discovered two isotopes of dubnium produced by the bombardment of ^{243}Am by ^{22}Ne ions. However, the Lawrence Radiation Laboratory work did not confirm these findings (due to energy and decay differences). *See* TRANSURANIUM ELEMENTS. Albert Ghiarso

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1												13	14	15	16	17	18
1 H	2											5 B	6 C	7 N	8 O	9 F	2 He
3 Li	4 Be											13 Al	14 Si	15 P	16 S	17 Cl	10 Ne
11 Na	12 Mg	3	4	5	6	7	8	9	10	11	12	31 Ga	32 Ge	33 As	34 Se	35 Br	18 Ar
19 K	20 Ca	21 Sc	22 Ti	23 V	24 Cr	25 Mn	26 Fe	27 Co	28 Ni	29 Cu	30 Zn	49 In	50 Sn	51 Sb	52 Te	53 I	36 Kr
37 Rb	38 Sr	39 Y	40 Zr	41 Nb	42 Mo	43 Tc	44 Ru	45 Rh	46 Pd	47 Ag	48 Cd	81 Tl	82 Pb	83 Bi	84 Po	85 At	54 Xe
55 Cs	56 Ba	71 Lu	72 Hf	73 Ta	74 W	75 Re	76 Os	77 Ir	78 Pt	79 Au	80 Hg						86 Rn
87 Fr	88 Ra	103 Lr	104 Rf	105 Db	106 Sg	107 Bh	108 Hs	109 Mt	110 Ds	111 Rg	112	113					

lanthanide series	57 La	58 Ce	59 Pr	60 Nd	61 Pm	62 Sm	63 Eu	64 Gd	65 Tb	66 Dy	67 Ho	68 Er	69 Tm	70 Yb

actinide series	89 Ac	90 Th	91 Pa	92 U	93 Np	94 Pu	95 Am	96 Cm	97 Bk	98 Cf	99 Es	100 Fm	101 Md	102 No

Ducted fan

A propeller or multibladed fan inside a coaxial duct or cowling, also called a ducted propeller or a shrouded propeller, although in a shrouded propeller the ring is usually attached to the propeller tips and rotates. The duct serves to protect the fan blades from adjacent objects and to protect objects from the revolving blades, but more importantly, the duct prevents radial flow of the fluid at the blade tips. Fan efficiency remains high over a wider speed range with a properly shaped duct than without. However, fan efficiency is sensitive to duct shape at off-center design conditions. Without a well-rounded inlet lip and a variable-area exit, off-center performance may be worse than without the duct. *See* VENTILATION.

With a duct, static thrust for a given power input is higher than without one (**Fig. 1**). For this reason, propellers of vertical takeoff and landing aircraft may be ducted. At low speeds, a stator of radial airfoils downstream from the propeller or an oppositely turning coaxial propeller improve efficiency by converting slipstream rotation into axial velocity. The duct may also form a nozzle to further increase exit jet velocity. Airflow past the outer contour of the duct influences overall performance (**Fig. 2**). *See* VERTICAL TAKEOFF

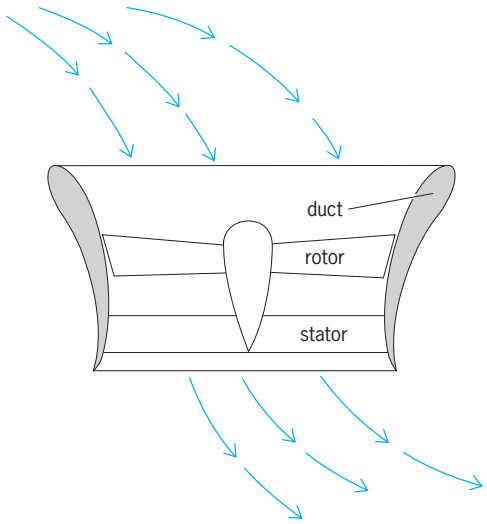

Fig. 1. Static thrust and operating speed range of fan is improved by a duct.

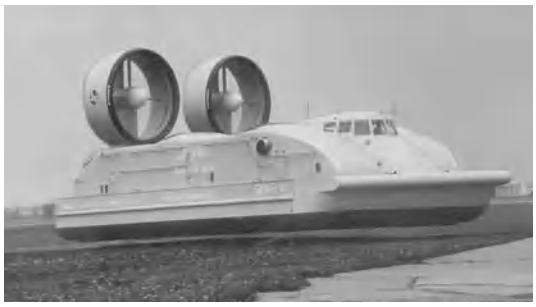

Fig. 2. Two ducted fans provide propulsion for Bell-Navy Hydroskimmer. (*Bell Aerosystems Co.*)

AND LANDING (VTOL).

Ducted fans are used in axial-flow blowers or compressors of several stages for turbine engines. In such applications, solidities are higher than for usual propellers, and stators or countervanes are usual in each stage. A ducted fan engine is a gas turbine arranged to move a larger mass of air than passes through the turbine, the additional air leaving at lower exit velocity and hence higher jet propulsion efficiency for moderate-speed aircraft than obtainable with a simple turbojet. *See* AIRCRAFT PROPULSION; CONVERTIPLANE; FAN; GAS TURBINE; TURBOFAN; TURBOJET.

Frank H. Rockett

Dumortierite

A nesosilicate mineral having the composition $Al(BO)_3(SiO_4)_3O_3$. Dumortierite crystallizes in the orthorhombic system but well-formed crystals are rare; the mineral usually occurs in parallel or radiating fibrous aggregates. There is one direction of poor cleavage. The hardness is 7 on Mohs scale, and the specific gravity is 3.26–3.36. The mineral has a vitreous luster and a color that varies not only from one locality to another but in a single specimen. It may be pink, green, blue, or violet. Dumortierite is found in schists and gneisses and more rarely in pegmatites. In the United States it occurs at Dehesa, California, and at Rochester, Nevada, where it has been mined for the manufacture of high-grade porcelain. *See* SILICATE MINERALS.

Cornelius S. Hurlbut, Jr.

Dune

Mobile accumulation of sand-sized material that occurs along shorelines and in deserts because of wind action. Dunes are typically located in areas where winds decelerate and undergo decreases in sand-carrying capacity. Dunefields are composed of rhythmically spaced mounds of sand that range from about 3 ft (1 m) to more than 650 ft (200 m) in height and may be spaced as much as 5000 ft (1.5 km) apart. Smaller accumulations of windblown sand, typically ranging in height from $^1/_4$ to $^1/_2$ in. (5 to 15 mm) and in wavelength from 3 to 5 in. (7 to 12 cm), are known as wind ripples. Dunes and ripples are two distinctly different features. The lack of intermediate forms shows that ripples do not grow into dunes. Ripples commonly are superimposed upon dunes, typically covering the entire upwind (stoss) surface and much of the downwind (leeward) surface as well.

Grain size and shape. Sand, which includes any material between 0.002 and 0.08 in. ($^1/_{16}$ and 2 mm) in diameter, moves across and between dunes primarily by the process of saltation (**Fig. 1**). Individual saltating sand grains move at nearly the same velocity as the wind and rarely reach more than 6 ft (2 m) above ground surface. When a fast-moving grain "splashes" into the underlying sand, it ejects other grains into the wind, thus perpetuating the transport process. The high-velocity, grain-to-grain impacts that take place during saltation transport produce well-rounded sand particles with distinctive surface pits that can be observed with the scanning electron microscope. Particles ranging from 0.08 to 0.2 in. (2 to 5 mm) in diameter creep slowly downwind as they are hit by saltating grains; these grains make up a small proportion of some dunes. Larger grains cannot be moved by normal winds and are therefore generally lacking in dune deposits. When particles finer than sand size are set into motion by the wind, they travel in suspension within high dust clouds. For this reason, silt and clay particles are not abundant in dunes but are deposited far downwind as broad sheets of loess. *See* LOESS; SAND.

Materials. Virtually any kind of sand-sized material can accumulate as dunes. The majority of dunes are composed of quartz, an abundant and durable mineral released during weathering of granite or sandstone. Dunes along subtropical shorelines, however, are commonly composed of grains of calcium carbonate derived in part from the breakdown of shells and coral. Along the margins of seasonally dry lakes, dunes may be composed of gypsum (White Sands, New Mexico) or sand-sized aggregates of clay minerals (Laguna Madre, Texas). *See* CLAY MINERALS; GYPSUM; QUARTZ.

Dynamics. The leeward side of most dunes is partly composed of a slip face, that is, a slope at the angle of repose. For dry sand, this angle is approximately 33°.

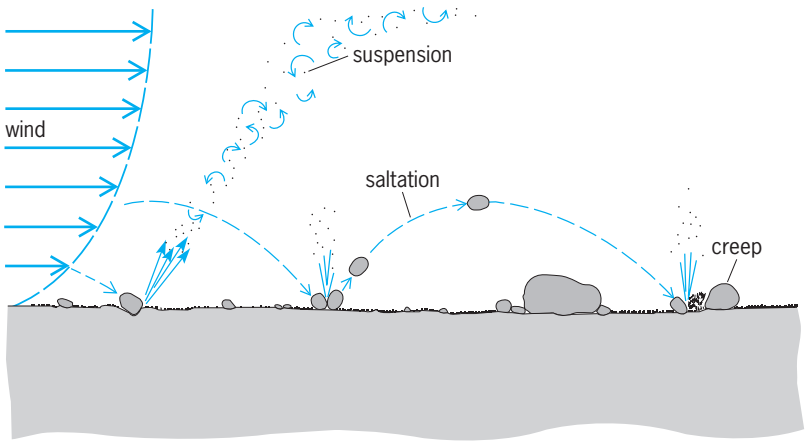

Fig. 1. Wind moves sand-sized particles by the process of saltation. Coarser material may creep or be immobile; finer material travels in suspension. Dunes are relatively free of materials coarser or finer than sand. (*After T. A. Mutch et al., Geology of Mars, Princeton University Press, 1976*)

Fig. 2. Tongue-shaped sand avalanches on the slip face of a barchanoid ridge dune on the eastern coast of Australia. (*Courtesy of D. B. Loope*)

When additional sand is deposited at the top of such a slope, tonguelike masses of sand avalanche to the base of the slope (**Fig. 2**). The dune migrates downwind as material is removed from the gently sloping stoss side of the dune and deposited by avalanches along the slip face. Much of the sand on the leeward side of the dune is later reworked by side winds into wind ripple deposits. Because the coarsest grains preferentially accumulate at the crests of wind ripples, the layering in wind ripple deposits is distinctive and relatively easily recognized—each thin layer is coarser at its top than at its base.

Sand deposited by avalanching is very loosely packed; up to 50% of its volume is occupied by air space. Wind ripple deposits are much more closely packed. A person with a knowledge of dune dynamics can choose a path which goes over firm wind ripple deposits and avoids loose avalanche deposits. Those less knowledgeable are likely to become bogged when their feet or tires sink deeply into the avalanche deposits.

Types. Dunes can be classified on the basis of their overall shape and number of slip faces. Three kinds of dunes exist with a single slip face; each forms in areas with a single dominant wind direction. Barchans are crescent-shaped dunes; their arms point down-wind. They develop in areas in which sand is in small supply. If more sand is available, barchans coalesce to form sinuous-crested dunes called barchanoid ridges. Transverse dunes with straight crests develop in areas of abundant sand supply. The axis of each of these dune types is oriented at right angles to the dominant wind, and the dunes migrate rapidly relative to other dune types. The migration rate of individual dunes is quite variable, but in general, the larger the dune, the slower the migration rate. In the Mojave Desert of southern California, barchans having slip faces 30 ft (10 m) long migrate about 50 ft (15 m) per year.

Dunes having more than one slip face develop in areas with more complex wind regimes. Linear dunes, sometimes called longitudinal or self dunes, possess two slip faces which meet along a greatly elongated, sharp crest (**Fig. 3**). Some linear dunes in Saudi Arabia reach lengths of 120 mi (190 km). Recent experimental evidence has shown that linear dunes are the result of bidirectional winds that differ in direction by more than 90°. The trend of these dunes is controlled by wind direction, strength, and duration, but the nature of the wind regime cannot be deduced from a knowledge of dune trends. Star dunes bear many slip faces and consist of a central, peaked mound from which several ridges radiate. Because they do not migrate appreciably, they grow in height as sand is delivered to them, some reaching 1000 ft (300 m).

Plant growth appears to be important to the growth and maintenance of two types of dunes. Coppice dunes are small mounds of sand that are formed by the wind-baffling and sand-trapping action of desert plants. The crescentic shape of parabolic dunes gives them a superficial resemblance to barchan dunes, but their arms point upward. Plants commonly colonize and anchor only the edges of a dune, leaving the body of the dune free to migrate. The retarded migration rate of the dune margin leads to the formation of the trailing arms of a parabolic dune. *See* DUNE VEGETATION.

Other dune types are dependent on special topographic situations for their formation. Climbing dunes develop on the upwind side of mountains or cliffs; falling dunes are formed at the sheltered, downward margin of similar features.

Distribution in space and time. Land plants are sufficiently abundant in most areas of the Earth to protect soil or sediment particles from wind erosion. It is only in areas that lack vegetation, such as along beaches and in deserts, that sufficient sand is moved to produce dunes. Unlike Earth, Mars lacks a vegetative cover, and dunes are abundant over a large portion of its surface.

Although all deserts of the world possess dunes, deserts occupy less than 1% of the arid portions of the southwestern United States. In the deserts of northern Africa, Saudi Arabia, and Australia, dunes cover approximately 30% of the ground surface. Coastal dunes are widespread except in the tropics, where heavy rainfall, dense vegetation, and relatively light winds severely limit their formation.

Deeply weathered, tree-covered dunes in South America and Africa indicate that during the last ice age the tropics were much more arid than today. Cooler global temperatures during this period

Fig. 3. Linear dune, Imperial County, California.

apparently decreased rainfall and increased wind strength, yielding conditions favorable for the expansion of dune fields.

Before the evolution and diversification of land plants on Earth, dunes may have been more widespread than today. Sandstones representing lithified, ancient dunefields have been identified on nearly all continents and are especially numerous and thick in the southwestern United States. In southern Utah, the Navajo Sandstone of Early Jurassic age (approximately 200 million years before present) reaches a thickness of over 2000 ft (600 m). The long, steeply inclined layers, or cross-beds, that were deposited during dune migration are characteristic features of these sandstones. Because dunelike mounds of sand can also form and migrate underwater, the most significant diagnostic criterion for recognizing eolian (windblown) sandstones is the presence of the distinctive wind ripple deposits. Eolian sandstones commonly contain animal trackways and traces of plant roots, and they are devoid of marine fossils larger than sand size. Typically they are found interbedded with ancient coastal or river deposits and may form important reservoirs for oil, natural gas, and fresh water. *See* EOLIAN LANDFORMS; SANDSTONE.

D. B. Loope

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Dune vegetation

Plants occupying sand dunes and the slacks, or swales, and flats between them. The density and diversity of dune vegetation are greater on coastal dunes than on desert dunes. *See* DUNE.

Zonation. A zonation pattern is evident in the vegetation of the coastal dunes (**Fig. 1**). A wrack, or debris, line occurs at the upper limit of the beach. Seeds caught in decaying plant material and other debris washed in on the high tides germinate here and trap windblown sand, initiating the formation of a dune. Small mounds formed in this way may eventually become foredunes as more sand accumulates and as the foredune grasses invade these mounds. Decomposing plant material (detritus) washed in on the tides provides additional nourishment for the young plants.

The foredunes, also called the primary dunes, are those closest to the water and lie behind the wrack line. The plants on these dunes, mostly grasses, are tolerant to sea spray, high winds, and sand accretion. By a system of underground stems called rhizomes, they overcome burial by sand and spread throughout the dune with new shoots arising from buds on the rhizomes. The roots and rhizomes of these dune grasses are important in stabilizing the dune sand and preventing wind erosion. The foredune plants participate in dune formation; by slowing the wind, they favor sand deposition.

Behind the primary dunes are the secondary dunes, sometimes called the dune field. In this more favorable environment, the vegetation is denser and more diverse; the foredunes block sea spray and reduce wind velocity. Here are found plants adapted to dry land and those tolerant of flooding, which often occurs in the dune slacks. The dune slacks are the low areas between dunes and are frequently a result of a blowout, an area where sand has been blown away down to where the sand is moist and close to the water table. Plants typical of wetlands often vegetate these areas. Shrub communities also inhabit the dune field and often form dense patches of vegetation.

A maritime forest may be found behind the secondary dunes. In coastal barrier beach or island locations, a salt marsh adjacent to a bay or sound may lie behind the forest.

Plant communities. Not all the plant species occupying the sand dunes of the United States can be described here. Although the names of the plants vary from coast to coast, their functions and adaptations are essentially the same. Sea rocket (*Cakile maritima*), probably the most widely distributed wrack-line beach plant, can be found in many coastal regions. It is an annual plant that initiates dune formation by trapping sand and debris.

The foredune species are primarily responsible for forming and stabilizing the seaward-facing row of

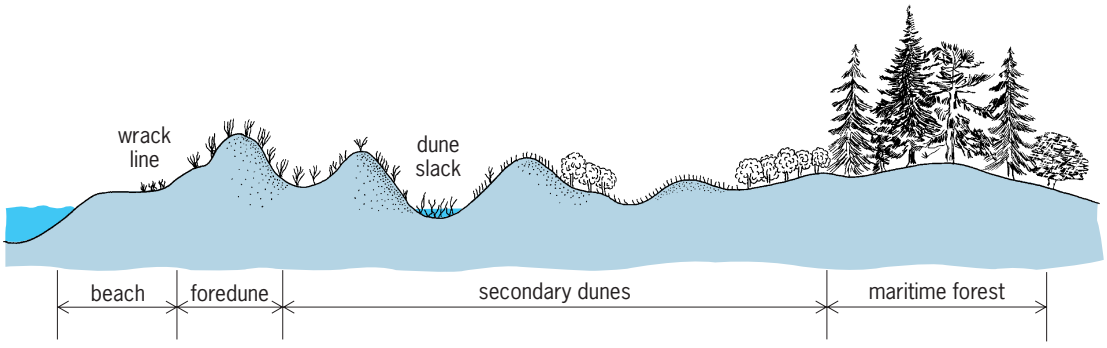

Fig. 1. A coastal dune profile.

Fig. 2. Sea oats are abundant on the foredunes of the Atlantic coast of the southeastern United States. Wrack, decomposing plant material, can be seen on the beach in the zone seaward of the sea oats.

dunes. American beachgrass (*Ammophila breviligulata*) is the dominant pioneer plant on the dunes of the North Atlantic and mid-Atlantic coast. This plant and prairie sandreed (*Calamovilla longifolia*) dominate the dunes along the shores of the Great Lakes. Sea oats (*Uniola paniculata*) dominates the foredunes of the South Atlantic coast (**Fig. 2**) and, with bitter panicum (*Panicum amarum*), is the pioneer species on the Gulf of Mexico dunes. On the northern west coast, European beachgrass (*Ammophila arenaria*) and American dunegrass (*Elymus mollis*) dominate. On the southern west coast, species of *Carpobrotus, Artemisia, Ambrosia,* and *Abrona* are most abundant.

In the secondary dunes the diversity increases. On the Atlantic coast, shrub communities include bayberry (*Myrica pensylvanica*), wax myrtle (*Myrica cerifera*), beach plum (*Prunus maritima*), and poison ivy (*Rhus radicans*). In the south, saw palmetto (*Serenoa repens*) and yucca (*Yucca filamentosa*) are common. On the Pacific coast, the shrub community includes salal (*Gaultheria shallon*), evergreen huckleberry (*Vaccinium ovatum*), chaparral broom (*Baccharis pilularis*), California figwort (*Schrophularia californica*), hooker willow (*Salix hookeriana*), and Pacific wax myrtle (*Myrica californica*). Seaside goldenrod (*Solidago sempervirens*), running beach grass (*Panicum amarum*), sand bur (*Cenchrus tribuloides*), beach heather (*Hudsonia tomentosa*), and beach pea (*Lathyrus maritimus*) are common on Atlantic and Gulf coast dunes. On the Pacific coast, common herbaceous plants are red fescue (*Festuca rubra*), dune goldenrod (*Solidago spathulata*), beach sagewort (*Artemisia pycnocephala*), and seashore lupine (*Lupinus littoralis*). On the dunes of the Great Lakes, blue joint (*Calamagrostis canadensis*), beach bea (*Lathyrus maritimus*), tansy (*Tanacetom huronense*), scouring rush (*Equisetum hyemale*), and beach heather (*Hudsonia tomentosa*) are common. Dune slacks of the Atlantic coast contain sedges such as American three-square (*Scirpus americanus*), rushes (*Juncus*), and grasses such as salt grass (*Distichlis spicata*). In the Pacific dune slacks, Douglas's spirea (*Spiraea douglasii*), slough sedge (*Carex obnupta*), Pacific silverweed (*Potentilla pacifica*), and common cattail (*Typha latifolia*) are common.

Forest species of the Atlantic and Gulf coast dunes include pitch pine (*Pinus rigida*), loblolly pine (*Pinus taeda*), sassafras (*Sassafras albidum*), wild black cherry (*Prunus serotina*), live oak (*Quercus virginiana*), and red cedar (*Juniperus virginiana*). Pacific dune forests include lodgepole pine (*Pinus contorta*), Douglas-fir (*Pseudotsuga menziesii*), western hemlock (*Tsuga heterophylla*), and sitka spruce (*Picia sitchensis*). Balsam poplar (*Populus balsamifera*) is the most common forest species of the Great Lakes dunes.

Ecology. Plants growing on sand dunes are adapted to the environment. The plants closest to the sea are usually the most tolerant of salt spray. The plants in the wrack line must be tolerant to salinity, wind, and burial by sand. Succulence helps sea rocket to tolerate the salty environment of the beach. It stores water in its fleshy stems and leaves, helping to dilute the salt taken up by its roots. Its tough cuticle withstands sand abrasion. *See* PLANTS, SALINE ENVIRONMENTS OF.

The foredune plants must be tolerant of sand burial, sea spray, and a nutrient-poor substrate. American beachgrass, for example, grows more vigorously in a location of accreting sand than in a more stable area. Its rhizomes, like those of other foredune grasses, can grow and spread rapidly as new sand is added to a dune. Once buried, these plants send a vertical rhizome to the surface, where a new shoot develops. Furthermore, some of these plants, such as American beachgrass, have specialized bacterium named *Azotobacter* associated with their roots, which fix atmospheric nitrogen into a form usable by the plant. These bacteria enable the plants to grow in an otherwise nitrogen-poor environment. Young roots secrete carbohydrates which may nourish the nitrogen-fixing bacteria. *See* NITROGEN FIXATION.

The plants in the dune slacks have morphological and physiological adaptations for growth in flooded areas. For example, the sedge American three-square contains large air spaces (aerenchyma) in its stems and roots, which provide an oxygen pathway from the shoots above the water to the oxygen-deprived roots in the flooded soil. When these plants are flooded, they increase their production of the hormone ethylene, which may stimulate the production of aerenchymatous tissue.

Management. Sand dunes form a natural defense against the sea; vegetation holds the sand in place and favors its accretion. Although dune vegetation is very tolerant of wind, windblown sand, and salt, it is sensitive to many human activities. Coastal areas are highly populated, especially in summer. The protection of these valuable ecosystems requires good management practices. Access to the beach through

Fig. 3. After one growing season, this planting of American beachgrass is already helping to stabilize this dune. Snow fence was used to trap and hold sand prior to planting.

the dunes is controlled, and damaged areas are restored with the help of porous fencing, such as snow fence, to trap and accumulate sand in eroded areas. Such fencing can increase the height and width of dunes. The new sand is planted with dune vegetation (**Fig. 3**) growing nearby or in commercial nurseries. These practices have been very successful.

Human disturbances include pedestrian and vehicular traffic, which can cause serious damage. In one typical area that was walked over less than 200 times in a year, more than half of the vegetation disappeared, leaving barren ground.

Off-road vehicular traffic is another serious threat. At the wrack line, it prevents the formation of new dunes by destroying organic matter and causing the loss of nutrients, water, and the seedlings themselves. Traffic behind the foredunes is equally destructive. Plants such as beach heather can be completely destroyed, needing years to recover. Tire tracks may remain visible for many seasons. Laws help but may be difficult to enforce. The complete abolition of beach vehicles may become necessary.

The protection of vegetation on the remaining dunes is important for several reasons. Dunes not stabilized by vegetation are highly susceptible to wind and water erosion and threatened with destruction. Their disappearance means a loss of nesting areas for many seabirds, other wildlife habitat, and valuable recreational areas. It would also leave settled areas, now protected by coastal dunes, open to erosion by wind and sea. *See* EROSION.

Denise M. Seliskar

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Dunite

An ultramafic igneous rock composed of at least 90% olivine. Important accessory minerals (abundances usually less than 1%) found in different occurrences of dunite include chromian spinel, low- or high-calcium pyroxenes (enstatite and diopside), and plagioclase. If these minerals constitute greater than 10% of the rock, it is called chromitite, peridotite, or troctolite, respectively. Low-temperature alteration (less than 750°F or 400°C) causes hydration of olivine to the mineral serpentine and, where extensive, may transform dunite to the metamorphic rock serpentinite. Dunite was first described at Dun Mountain, New Zealand. The dun color is a characteristic feature of the weathered rock. *See* CHROMITE; OLIVINE; PERIDOTITE; PYROXENE; SERPENTINITE.

Composition. Dunite is ultrabasic in composition, meaning that it is low in silica compared to most crustal rocks (<45% silica) and is generally very high in magnesium (up to 54%). In some occurrences, notably in continental layered intrusions, however, the olivine may be very iron rich (variety hortonolite), and the rock can contain as much as 40% iron. Dunite is notably poor in alumina, soda, and lime; and while it may be relatively nickel rich, many other critical trace elements are nearly missing. Thus, weathering of dunite forms soils to which most terrestrial plants are poorly adapted, and such soils often host unusual plant species such as the carnivorous cobra lily and the miniature rhododendron *Kalmiopsis leachiana* found in the coast ranges of Oregon and California.

Texture and structure. Olivine crystals in some dunites can often be seen originally to have had equant shapes where the original crystal faces are in contact with intergranular plagioclase and pyroxene. Commonly, however, the olivine crystals have grown together during crystallization from a magma, or have undergone annealing recrystallization following high-temperature deformation (as commonly occurs during solid-state mantle flow) to form granular aggregates of irregularly shaped equidimensional crystals. *See* MAGMA.

Origin. In layered and southeast Alaskan-type intrusions, dunite is formed by fractional crystallization of basic magmas, whereby relatively dense olivine crystallizes and settles out of the melt, often with minor chromite, to form nearly pure pockets and layers. Dunite also forms during late upward transport by porous flow of melt in pipes through the nearly crystalline intrusions, whereby the original minerals are replaced by olivine because of reaction with the melt. Metasomatic reaction between late hydrous fluids and peridotites may also occur to produce some crosscutting dunites. Elsewhere, dunites may have formed by flow differentiation of olivine crystals in

basic dikes during melt transport along cracks and fissures in the Earth. *See* METASOMATISM; PLUTON.

In the mantle, dunite may form by precipitation of olivine from magmas migrating upward through the Earth, or by reaction of melts formed at high pressures with the mantle during upward transport along cracks or during focused porous flow where the melt is moving too fast to maintain simple equilibrium with the mantle. Thus, the occurrence of dunite is believed to mark the location of conduits where melt was transported through the Earth. *See* EARTH.

Occurrence. Dunite is found in continental associations as layers near the base of the large continental layered intrusions (fossil magma chambers crystallized deep within the crust), such as the Bushveld Complex of South Africa. It also occurs as crosscutting vertical pipes in these and in the concentric basic intrusions of the southeast Alaskan type found associated with the basement of volcanic arc terrains.

The most important occurrence of dunite is in its association with rocks believed to have come from the Earth's mantle (principally peridotite) where it marks the location of former conduits through which magmas have been transported out of the Earth to the crust. In the oceans, it is infrequently dredged from the great oceanic transform faults cutting the ocean ridges, exposed in tectonic windows in disrupted and uplifted ocean crust in association with serpentinized mantle peridotite. In 1993 and 1994, however, important dunite bodies crosscutting tectonically exposed mantle sections on the walls and floors of rift valleys have been found in the central Atlantic and eastern Pacific oceans far from transform faults. The suggestion is that the flow of magma out of the mantle upwelling beneath ocean ridges is not uniform, and is therefore important in controlling the formation and structure of the two-thirds of the Earth's crust that is formed in the oceans. *See* MID-OCEANIC RIDGE; RIFT VALLEY; TRANSFORM FAULT.

Dunite is also found as lenses, pods, and veins in alpine-type orogenic belts, where it is found crosscutting subcontinental mantle peridotites, in xenoliths of broken rock carried up to the surface by magmas erupting from the mantle, and in mantle rocks underlying ophiolite complexes. The latter are fossil sections of ocean crust and mantle, tectonically emplaced on-land during plate-tectonic collisions along continental margins and island arcs, and believed to have originally formed in island arcs and marginal seas. *See* CONTINENTAL MARGIN; EARTH CRUST; OPHIOLITE; PLATE TECTONICS; XENOLITH.

Uses. Dunite has been extensively quarried as a building stone because of its unusual pale water-green color, but is little used now because of its high susceptibility to acid rain. Dunite is also used for refractory materials for furnaces. Olivine in dunite is occasionally of gem quality, as on Zabargad Island in the Red Sea. Dunite is also the principal host of deposits of the mineral chromite. The hortonolite-dunite pipes found in the Bushveld Complex of South Africa have also been mined for platinum-group minerals. *See* GEM; IGNEOUS ROCKS; PETROLOGY; REFRACTORY.

Henry J. B. Dick

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Dust and mist collection

The physical separation and removal of particles, either solid or liquid, from a gas in which they are suspended. Such separation is required for one or more of the following purposes: (1) to collect a product which has been processed or handled in gas suspension, as in spray-drying or pneumatic conveying; (2) to recover a valuable product inadvertently mixed with processing gases, as in kiln or smelter exhausts; (3) to eliminate a nuisance, as in fly-ash removal; (4) to reduce equipment maintenance, as in engine intake air filters; (5) to eliminate a health, fire, explosion, or safety hazard, as in bagging operations or nuclear separations plant ventilation air; and (6) to improve product quality, as in cleaning of air used in processing pharmaceutical or photographic products. Achievement of these objectives involves primarily gas-handling equipment, but the design must be concerned with the properties and relative amounts of the suspended particles as well as with those of the gas being handled.

All particle collection systems depend upon subjecting the suspended particles to some force which will drive them mechanically to a collecting surface. The known mechanisms by which such deposition can occur may be classed as gravitational, inertial, physical or barrier, electrostatic, molecular or diffusional, and thermal or radiant. There are also mechanisms which can be used to modify the properties of the particles or the gas to increase the effectiveness of the deposition mechanisms. For example, the effective size of particles may be increased by condensing water vapor upon them or by flocculating particles through the action of a sonic vibration. Usually, larger particles simplify the control problem. To function successfully, any collection device must have an adequate means for continuously or periodically removing collected material from the equipment.

Devices for control of particulate material may be considered, by structural or application similarities, in eight categories, which follow.

Gravity settling chamber. In this, the simplest type of device but not necessarily the least expensive, the velocity of the gas is reduced to permit particles to settle out under the action of gravity. Normally, settling chambers are useful for removing particles larger than 50 micrometers in diameter, although

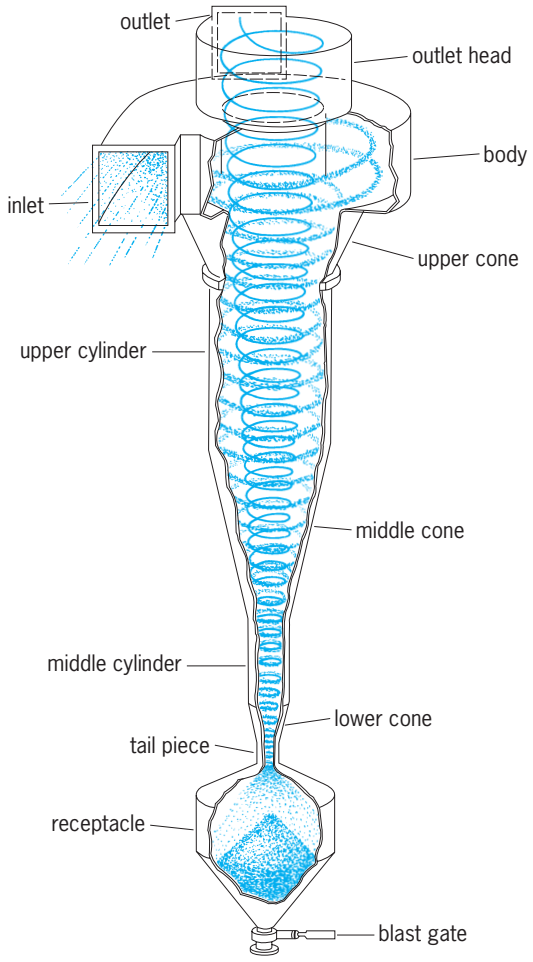

Fig. 1. Cyclone dust separator, an inertial device. (*American Standard Industrial Division*)

with special configurations they may be used to remove particles as small as 10 μm.

Inertial device. The basis of this type of device is that the particles have greater inertia than the gas. The cyclone separator, typical of this type of equipment, is one of the most widely used and least expensive types of dust collector. In a cyclone, the gas usually enters a conical or cylindrical chamber tangentially and leaves axially. Because of the change of direction, the particles are flung to the outer wall, from which they slide into a receiving bin or into a conveyor, while the gases whirl around to the central exit port (**Fig. 1**). A large variety of configurations is available. For large air-handling capacities, an arrangement of multiple small-diameter units in parallel is often used to attain high collection efficiencies and to permit lower headroom requirements than a single unit would.

Mechanical inertial units are similar to cyclones except that the rotational motion of the gas is induced by the action of a rotating member. Some such units are designed to act as fans in addition to their dust-collecting function. There are also a wide variety of other units; many are called impingement separators. Most separators that are used to remove entrained liquids from steam or compressed air fall into this category.

Packed bed. A particle-laden gas stream may be cleaned by passing it through a bed or layer of packing composed of granular materials such as sand, coke, gravel, and ceramic rings, or fibrous materials such as glass wool, steel wool, and textile staples. Depending on the application, the bed depth may range from a fraction of an inch to several feet. Coarse packings, which are used at relatively high throughput rates (1–15 ft/s or 0.3–4.5 m/s superficial velocity) to remove large particles, rely primarily on the inertial mechanism for their separating action. Fine packings, operated at lower throughput rates (1–5 ft/min or 0.3–15 m/min superficial velocity) to remove relatively small suspended particles, usually depend on a variety of deposition mechanisms for their separating effect. Packed beds, because of a gradual plugging caused by particle accumulation, are usually limited in use to collecting particles present in the gas at low concentration, unless some provision is made for removing the dust—for example, by periodic or continuous withdrawal of part of the packing for cleaning. Depending on the application and design, the collection efficiencies of packed beds range widely (50–99.999%).

Cloth collector. In such a collector, also known as a bag filter, the dust-laden gas is passed through a woven or felted fabric upon which the gradual deposition of dust forms a precoat, which then serves as a filter for the subsequent dust. These units are analogous to those used in liquid filtration and represent a special type of packed bed. Because the dust accumulates continuously, the resistance to gas flow gradually increases. The cloth must, therefore, be periodically vibrated or flexed, or back-flushed with a stream or pulse of air to dislodge accumulated dust (**Fig. 2**). A wide variety of filter media is

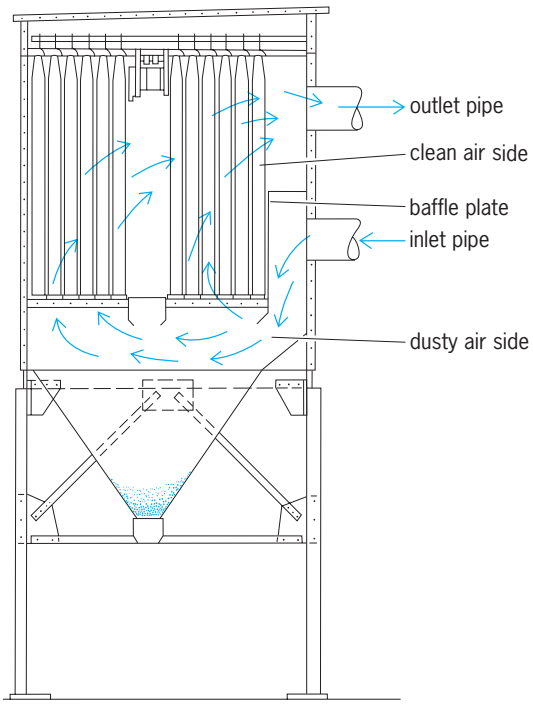

Fig. 2. Cloth collector. (*Wheelabrator Corp.*)

available. Cotton or wool sateen or felts are usually used for temperatures below 212°F (100°C). Some of the synthetic fibers may be used at temperatures up to 500°F (260°C). Glass and asbestos or combinations thereof have been used for temperatures up to 650°F (343°C). For special high-temperature applications, metallic screens and porous ceramics or stainless steel have been employed. Collection efficiencies of over 98% are attained readily with cloth collectors, even with very fine dusts.

Scrubber. A scrubber uses a liquid, usually water, to assist in the particulate collection process. An extremely wide variety of equipment is available, ranging from simple modifications of corresponding dry units to permit liquid addition, to devices specifically designed for wet operation only (**Fig. 3**). When properly designed for a given application, scrubbers can give very high collection efficiency, although the mere addition of water to a gas stream is not necessarily effective. For a given application, collection efficiency is primarily a function of the amount of power supplied to the gas stream, in the form of gas pressure drop, water pressure, or mechanical energy. With scrubbers it is important that proper attention be given to liquid entrainment separation in order to avoid a spray nuisance. Consideration must also be given to the liquid-sludge disposal problem.

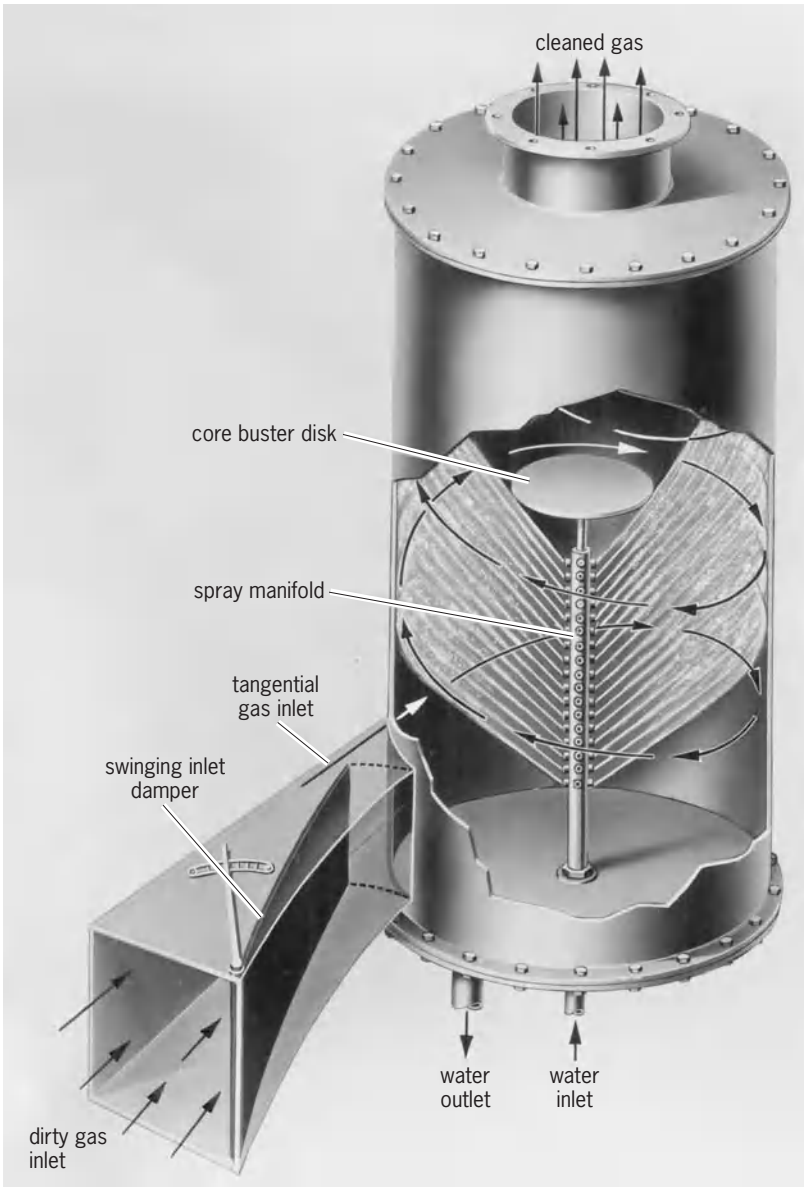

Fig. 3. Cyclonic liquid scrubber. (*Chemical Construction Corp.*)

Electrostatic precipitator. Particles may be charged electrically by a corona discharge and caused to migrate to a collecting surface. The single-stage unit, which is commonly known as a Cottrell precipitator, and in which the charging and collecting proceed simultaneously, is the type generally used for industrial or process applications. These units normally employ direct current at voltages ranging from 30,000 to 100,000 V. The two-stage unit, in which charging and collection are carried out successively, is commonly used for air conditioning applications. These units also employ direct current, ranging from 5000 to 13,000 V, and involve close internal clearances (0.25–0.5 in. or 6–12 mm). Electrical precipitators are capable of high collection efficiency of fine particles.

In dry precipitators, collected dust is dislodged by intermittent rapping or vibration of the collection plates with possible dust reentrainment. Another problem arises from high dust resistivity, which causes collected dust to be reentrained by back-ionization because the dust particles are not discharged electrically upon collection. The reentrainment of collected material as flocs in the exhaust gas, a phenomenon known as snowing, must be avoided to prevent a possible accentuated nuisance or vegetation damage problem.

In "wet" electrostatic precipitators, a film of water is allowed to flow over the collecting surface. This serves to minimize problems of dust reentrainment either by mechanical means or by the process of back-ionization. It also provides a means of removing collected dust. *See* ELECTROSTATIC PRECIPITATOR.

Air filter. This is a unit used to eliminate very small quantities of dust from large quantities of air, as in air conditioning applications. Although units in this class actually fall into one of the previous classes, they are given a special category because of wide usage and common special features. In this category are viscous-coated fiber-mat filters and dry filters. These are actually a form of packed bed and are frequently known as unit filters. The domestic furnace filter is an example of a viscous-coated unit filter. Automatic filters provided with continuous and automatic cleaning arrangements are available in both the viscous-coated and dry forms, as well as with electrostatic provisions. *See* AIR FILTER.

Miscellaneous equipment. Electrostatic fields have been used with packed beds, scrubbers, and air filters to improve collection efficiency, and with cloth collectors to dislodge dust from the cloth. Acoustic or sonic vibrations imparted to a gas stream cause particulates to collide and flocculate, forming larger particles that are more readily collected in conventional apparatus. This principle has been employed,

but has had extremely limited application because of economic and other practical considerations. In thermal precipitation, suspended particles are caused to migrate toward a cold surface or away from a heated surface by the action of a temperature gradient in the gas stream. This principle has found extensive use in atmospheric sampling work. *See* AIR POLLUTION; CHEMICAL SEPARATION TECHNIQUES; MECHANICAL SEPARATION TECHNIQUES; SMOKE; UNIT OPERATIONS.

Charles E. Lapple

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Dust storm

A strong, turbulent wind bearing large clouds of dust. The suspended dust consists predominantly of mineral particles with diameters less than 100 micrometers. Sand storms differ from dust storms in that the particles are much larger and have high settling speeds. When the wind and turbulence weaken, the larger particles settle out rapidly, leaving behind a cloud of dust with diameters mainly under 20 μm. Wind systems can carry these clouds many thousands of miles, in some cases across oceans.

Sources of dust. Dust storms can occur where the soils are loosely consolidated and poorly protected by vegetation. Soil moisture is critically important, since water binds the soil particles together and supports the growth of plants. Consequently, storms occur most frequently in arid and semiarid regions (**Fig. 1**), with the maximum frequency in regions having an annual rainfall of 4–8 $\times$ 10^{-3} in. (0.1–0.2 mm). Hyperarid areas such as deep deserts are not particularly good sources, because the surface is often covered with rock and sand, neither of which contains an abundance of fine particles.

Marginal agricultural lands are especially susceptible to storms in the spring when the soils, loosened by plowing and overgrazing, are easily blown away. Extended periods of drought have the same effect, for example, the periodic dust bowl conditions in the Great Plains of the United States. *See* DROUGHT.

Most of the global dust storm activity is concentrated in a belt of arid lands that extends from the west coast of North Africa, through the Middle East and across Central Asia, almost to the Pacific coast. Many locations in this region experience dust storms 30 to 60 days a year.

Soil factors. In typical soils, dust generation begins when winds become strong enough to dislodge particles with diameters of several millimeters and larger (**Fig. 2**). These particles roll across the surface and bounce into the air, where they pick up speed from the wind. When they fall back to the surface, they strike other particles, blasting them into the air. This process is known as saltation. These new particles continue the process, producing an avalanche effect that rapidly sets the entire surface of the soil into motion. Particles larger than about 0.2 in. (5 mm) move in a shallow layer within a few feet of the ground. Smaller particles will be blasted into the air with the large particles; because of their smaller settling velocity, they will not fall back to the surface so rapidly. In this manner, a cloud of dust particles rapidly fills the turbulent atmospheric boundary layer. The depth of this boundary layer can vary greatly, depending on meteorological conditions.

Dust generation can be impeded in a number of ways. Vegetation reduces the force of the wind on the surface, and roots tend to hold the soil together. Soil cohesiveness is increased if the soil surface is minimally disturbed. Improper tilling and overgrazing

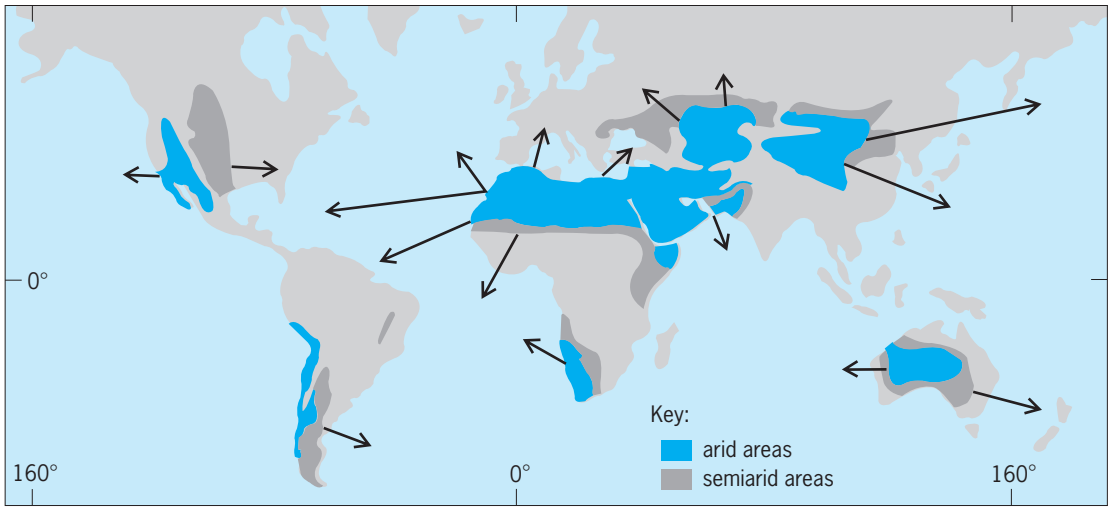

Fig. 1. Distribution of areas with high dust storm activity. The arrows indicate the major trajectories that dust storms follow. (*After K. Pye, Aeolian Dust and Dust Deposits, Academic Press, 1987*)

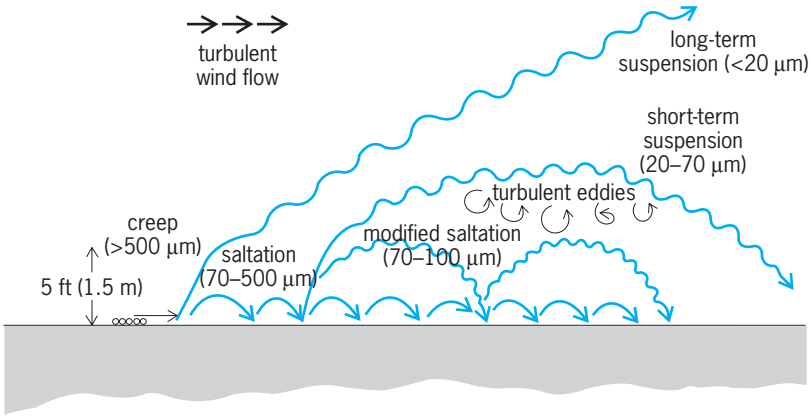

Fig. 2. Dust generation and transport near the soil surface. The bouncing dislodges both large particles, which continue the saltation process, and small particles, which can go into suspension. (*After K. Pye, Aeolian Dust and Dust Deposits, Academic Press, 1987*)

are prime factors in wind erosion. *See* AGRICULTURAL SOIL AND CROP PRACTICES; COVER CROPS; EROSION.

Meteorological factors. The wind speed required to generate dust varies widely according to soil characteristics. Typical values are 9–29 mi/h (4–13 m/s) for poorly sorted river sediments; 11–36 mi/h (5–16 m/s) for various desert environments including the deep Sahara; and 25–36 mi/h (11–16 m/s) for areas in Colorado and Arizona.

The area covered by the dust cloud and its height will depend on the meteorological conditions that generate the winds. Dust devils normally have a diameter of a few meters at most, and the height of the vortex is usually in the range of 10–300 ft (3–100 m). Although dust devils are small, they are commonplace, and in many areas they are probably responsible for much of the low-level dust generated under normal wind conditions.

In some areas, dust is commonly generated by cold downdrafts from cumulonimbus clouds. Known as haboobs, these generally last for 30 min to an hour, and they typically carry dust to heights of several hundred yards, at times to several thousand yards. The dust-generating area is confined to the region of the cloud and usually covers an area of tens to a few hundred square miles.

Major dust storms are associated with large-scale atmospheric features. The close positioning of high- and low-pressure centers produces a tight pressure gradient and strong winds. This situation is responsible for many dust events in the Middle East and for the Santa Ana in the western United States. Dust generated under these conditions is usually confined to the lower mile of the atmosphere. *See* AIR PRESSURE; WIND.

Cold fronts are responsible for truly large dust events that can be carried great distances. The vigorous winds associated with fronts are effective in generating dust over large areas. Moreover, the stronger atmospheric lifting ahead of the front can carry the dust to great altitudes, 10,000–23,000 ft (3–7 km). The high winds found at these altitudes can transport the dust thousands of miles in a few days. In China, storms of this type produce heavy dust over thousands of square miles and subsequently carry the dust over large areas of the North Pacific Ocean. The movement of these large dust clouds can be readily observed in satellite imagery. *See* AIR MASS; ATMOSPHERE; FRONT; SATELLITE METEOROLOGY.

Effects of dust. Dust plays a further role in weather in that mineral particles serve as condensation and freezing nuclei which are essential for the formation of clouds. Dust clouds typically contain particle concentrations of 10^{-7} to 10^{-4} oz/ft^3 (10^2 to 10^5 micrograms/m^3). These amounts can severely reduce visibility, causing problems for air, ship, and ground traffic over large areas. *See* CLOUD PHYSICS.

Dust can alter the distribution of solar energy in the Earth–atmosphere system, reducing the amount reaching the Earth's surface and producing heating in the troposphere. These effects can alter the dynamics of the atmosphere. Large amounts of dust as hypothesized in the nuclear winter scenarios would dramatically alter climate. *See* ATMOSPHERIC GENERAL CIRCULATION; CLIMATE MODIFICATION.

The impact of windblown dust on surfaces can produce large electrostatic charges that can affect the operation of electronic equipment. The inhalation of large quantities of dust can result in a variety of lung damage, the degree of damage depending on the composition of the dust (for example, asbestos). Organisms associated with the soil particles can also cause diseases, some of them quite serious.

J. M. Prospero

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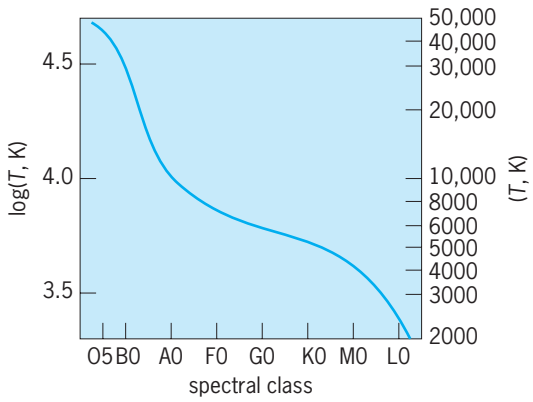
Dwarf star

A star whose state of evolution resembles that of the Sun. The term "dwarf star" derives from the early work of Ejnar Hertzsprung and Henry Norris Russell, who distinguished two kinds of stars, large ones called giants (or supergiants) and smaller ones called dwarfs. "Dwarf" is synonymous with "main sequence star" [Morgan-Keenan-Kellman (MKK) luminosity class V, where the luminosity classes are supergiants = I, bright giants = II, giants = III, subgiants = IV, and dwarfs = V], and implies not size but evolutionary condition. Dwarfs are stars that, like the Sun, fuse hydrogen into helium in their cores, the thermonuclear reactions providing energy and support. *See* GIANT STAR; HERTZSPRUNG-RUSSELL DIAGRAM.

Exact characteristics depend on chemical composition. Dwarfs of roughly solar composition (those in the galactic disk) range over the entire spectral sequence, while low-metal stars (the "subdwarfs" of the Galactic halo) occupy only the lower



Absolute temperatures of dwarf (main sequence) stars plotted against their spectral class. The absolute temperature *T*, measured in kelvins (K), is plotted on a logarithmic scale. Temperatures drop, following a sinuous curve, from a high of about 48,000 K (85,940°F) at the hot end to 2000 K (3140°F) at the low end. (*After J. B. Kaler, Cambridge Encyclopedia of Stars, Cambridge University Press, 2005*)

end of the main sequence from spectral class G on down (low abundances of metals, defined by astronomers as all elements more massive than helium, rendering the classes problematic). At the cool limit, which lies near the warm end of class L (about L2), dwarfs of solar composition have effective temperatures around 2000 K (3140°F), absolute visual magnitudes of +21, and bolometric luminosities somewhat under 10^{-4} that of the solar luminosity. At the hot limit (now extended to class O2 from the previous O3 limit), the values are, respectively, the order of 48,000 K (85,940°F), −6, and over 10^6 solar. The Sun, a G2 dwarf (that is, G2 V) falls in the middle at 5780 K (9945°F) with an absolute visual magnitude of +4.83 (see **illustration**). These properties are produced by a mass range from just under 0.08 solar masses for warm L dwarfs, below which full hydrogen fusion cannot be turned on (and the stars become brown dwarfs), to around 120 solar masses at classes O2–O3. On average, dwarf luminosities increase approximately as the 3.5 power of the mass, though the exponent varies from 2.7 at low mass, to 4 for solar mass stars, and back to 3 for stars of higher mass. The coolest dwarfs are small, with diameters at minimum under 0.1 solar, while O dwarfs can be quite large, over 20 times solar. The numbers of objects in the different spectral classes drop dramatically as one moves upward along the main sequence. From M through O, over 70% fall into class M, whereas 10% are in class G, and a mere 0.00004% are in class O. *See* BROWN DWARF; MAGNITUDE (ASTRONOMY); SUN.

Lower-mass dwarfs, below about 1.25 solar masses and spectral class F5, get their energy dominantly from the proton-proton chain, have low rotation speeds, and have convective envelopes with radiative cores (becoming entirely convective at the lowest masses). Higher-mass dwarfs get their energy dominantly from the carbon cycle, can be fast rotators, and have radiative envelopes and convective cores. At the high-mass end, above about 10 solar masses, the dwarf stage lasts 2–20 million years, and O stars turn into supergiants and core-collapse type II supernovae. At roughly class G8 and a mass of 0.85 solar, the lifetime equals the age of the Milky Way Galaxy; dwarfs between this limit and 10 solar masses become white dwarfs. *See* CARBON-NITROGEN-OXYGEN CYCLES; PROTON-PROTON CHAIN; SPECTRAL TYPE; STAR; STELLAR EVOLUTION; STELLAR ROTATION; SUPERGIANT STAR; SUPERNOVA; WHITE DWARF STAR. James B. Kaler

Dwarfism and gigantism

Underdevelopment and overdevelopment of the skeleton, respectively. Skeletal growth is a complex process and can be disturbed in many ways. For example, overstimulation by excessive growth-hormone production during childhood can produce gigantism. This is usually due to a pituitary tumor. Insufficient stimulation of skeletal growth, resulting from hormonal, metabolic, or nutritional disturbances, leads to reduced height with normal body proportions. The shortness of stature depends on the degree of the disturbance; the designation proportionate dwarfism is used in severe cases. Proportionate dwarfism may or may not be genetic. *See* PITUITARY GLAND.

Commonly, the term dwarfism is applied to short individuals with abnormal body proportions resulting from disturbances of bone growth itself. The bones of either limbs or trunk can be most affected. The disorders, called chondrodysplasias, number well over 100 and show a wide variation in clinical features: some appear at birth, others become evident in later childhood. The severity ranges from conditions in which infants die at birth to conditions compatible with normal life. In most instances, however, the abnormal skeleton creates medical problems. Joint deformities and osteoarthritis are common. Sometimes the skeletal abnormalities affect other organ systems, as in compression of the spinal cord at the base of the skull or in the lower back. Short ribs may reduce the size of the thoracic cavity, producing breathing difficulties. Other organ systems may also be affected, as in dysfunction of the kidneys or immune system. Intelligence is usually average. The specific condition is diagnosed on the basis of clinical features and skeletal x-rays.

Almost all of the chondrodsyplasias are inherited as mendelian or single-gene traits. However, because of reduced fertility and a high incidence of new mutations for certain traits, most individuals with these disorders are born to parents of average stature. Although it is known that the gene mutations disrupt endochondral ossification, which is responsible for skeletal development and linear bone growth, the mechanisms by which this occurs are poorly understood. *See* HUMAN GENETICS; SKELETAL SYSTEM DISORDERS. William A. Horton

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Dyadic

A mathematical abstraction corresponding to an expression of the type $\boldsymbol{\beta\gamma} + \boldsymbol{\delta\epsilon} + \cdots$, in which the elements (dyad symbols) consist of two vector symbols in juxtaposition without the intervention of either the dot ($\cdot$) or cross ($\times$). Essentially a dyad is an ordered pair of vectors subject to certain rules of operation. The first symbolic factor in a dyad (β in $\beta\gamma$, for example) is called the antecedent and the second the consequent.

The concept dyadic (J. W. Gibbs) is inherent in the formal structure of certain vectorial expressions such as $(\boldsymbol{\alpha} \cdot \mathrm{i})\mathrm{i} + (\boldsymbol{\alpha} \cdot \mathrm{j})\mathrm{j} + (\boldsymbol{\alpha} \cdot \mathfrak{k})\mathfrak{k}$ and $(\boldsymbol{\alpha} \cdot \boldsymbol{\beta})\boldsymbol{\gamma} + (\boldsymbol{\alpha} \cdot \boldsymbol{\delta})\boldsymbol{\epsilon}$. To reduce $(\boldsymbol{\alpha} \cdot \boldsymbol{\beta})\boldsymbol{\gamma} + (\boldsymbol{\alpha} \cdot \boldsymbol{\delta})\boldsymbol{\epsilon}$ to the form $\boldsymbol{\alpha} \cdot (\boldsymbol{\beta\gamma} + \boldsymbol{\delta\epsilon})$ it is sufficient to accept $\boldsymbol{\beta\gamma} + \boldsymbol{\delta\epsilon}$ as a mathematical entity and to agree that, whenever $\boldsymbol{\alpha} \cdot$ is applied from the left to $\boldsymbol{\beta\gamma} + \boldsymbol{\delta}\epsilon$, it is to be distributed to the antecedents, while the symbol $+$ is to become the plus of vector addition. Similarly, $(\boldsymbol{\beta\gamma} + \boldsymbol{\delta\epsilon}) \cdot \boldsymbol{\alpha} = \boldsymbol{\beta}(\boldsymbol{\gamma} \cdot \boldsymbol{\alpha}) + \boldsymbol{\delta}(\boldsymbol{\epsilon} \cdot \boldsymbol{\alpha})$ by distribution to the consequents.

Two dyadics Φ and Ψ are by definition equal if and only if $\Phi \cdot \boldsymbol{\alpha} = \Psi \cdot \boldsymbol{\alpha}$ identically in $\boldsymbol{\alpha}$. In particular, if $\boldsymbol{\beta} \neq 0$, $\boldsymbol{\gamma} \neq 0$, and $\boldsymbol{\beta\gamma} = \boldsymbol{\gamma\beta}$, then $\boldsymbol{\beta}(\boldsymbol{\gamma} \cdot \boldsymbol{\alpha}) = \boldsymbol{\gamma}(\boldsymbol{\beta} \cdot \boldsymbol{\alpha})$ and $\boldsymbol{\beta}$ and $\boldsymbol{\gamma}$ are parallel.

If the $\mathfrak{r}_a$ are the base vectors $\partial\mathfrak{r}/\partial x^a$ of a space defined by $\mathfrak{r} = \mathfrak{r}(x^1, \ldots, x^n)$ and $\mathfrak{r}^b = g^{bc}\mathfrak{r}_c$ and the quantities T^a_b are the components of a tensor of the type $(1, 1, w)$, then $T_b{}^a\mathfrak{r}_a\mathfrak{r}^b$ is an invariant. Consequently, tensors can be related to dyadics. *See* TENSOR ANALYSIS.

Homer V. Craig

Dye

A colored substance which imparts more or less permanent color to other materials. *See* DYEING.

Not all colored substances are dyes, however. If red iron rust is ground with white sugar, the resulting mixture has an overall reddish appearance. The sugar has been colored by pigmentation, and the iron rust has been used as a pigment. Examination of the mixture under a microscope shows distinct white and red particles, and a separation can be made by dissolving the sugar out of the mixture with water. If the red iron rust is added to white cloth in water, some of the red particles may cling to the cloth, but they can be removed by rubbing or by washing with soap. Some colored substances (usually organic chemical compounds) may be added to cloth in water, and after a period of soaking, usually accompanied by heat and agitation, the cloth will be colored; no separate colored particles can be seen under ordinary microscopic examination, and the color cannot be removed by washing, even with soap. The cloth has been dyed, and the colored substance is a dye (also called a dyestuff).

Customarily, colored water-insoluble substances are called pigments. Dyes are generally water-soluble, although some are soluble only during application, after which they become insoluble. *See* PIGMENT (MATERIAL).

The mechanism by which soluble colored substances enter the internal structure of fibers and there become fixed has been variously explained in terms of the physical and chemical concepts of the times when the explanations were given. It is said to be an adsorption phenomenon, a salt formation, a quasi-chemical union caused by hydrogen bonding, or an ether linkage, and in some cases it is considered to be a true solution effect. The end result, however, is that the dye has imparted a color (not necessarily that of the solid dye itself) to the fiber which is more or less resistant to washing or removal by similar mechanical operations. The dye is said to be fixed on and to have affinity, or substantivity, for the material it has colored. The material is designated as the substrate. If the color is quite resistant to washing and light, it is called a fast color; if the color is easily removed or fades quickly, it is a fugitive dye.

Because not all water-soluble colored substances are dyes, various attempts have been made to relate chemical constitution with color and substantivity. One of the earlier and still very useful explanations was given by O. N. Witt, who stated in 1876 that all colored organic compounds (called chromogens) contain certain unsaturated chromophoric groups which are responsible for the color, and if these compounds also contain certain auxochromic groups, they possess dyeing properties. Examples of chromophores are the groups —NO_2, —N=N—, and =CO, and of auxochromes, —NH_2, —OH. The auxochromes also influence hue according to their nature, number, and position on the chromogen molecule. An elaboration of this theory in 1888 by H. E. Armstrong regarded all chromophores as being quinoid (=R=) in structure. Many later studies have added explanations of the nature and variation of color in organic compounds, but the Witt theory provides a frame of reference for most dyes which is simple, practical, and satisfactory for all but the specialist in dye chemistry.

Dyestuffs may be classified in various ways, according to color (blue, red, and so on), origin (natural—from vegetable and animal matter—or synthetic), chemical structure, kinds of material to which they are applied, and method of application.

Color or hue. Commercial dyes are usually named to indicate the hue imparted to the dyed article. This color is not necessarily the same as that of the solid dyestuff, nor is it the same as the color of the dye solution from which the dyeing is made.

Origin. The dyes used earliest were natural dyestuffs (saffron, henna, cochineal, and logwood) derived from plants or animals. Most natural dyes are of the mordant type that requires a fixing agent. With the advent of the synthetic dyes, which are far more varied in shades and properties, and generally more brilliant, the production of the natural colors has decreased to a very small portion of the total market. Their greatest use is in the dyeing of leather.

Synthetic dyes were an early result of the development of chemistry as a science, and the first commercial production in 1857 in England of a synthetic organic chemical was that of the first commercially produced synthetic dye, mauve, discovered in 1856 by William Henry Perkin. The early history of organic

chemistry was primarily a record of investigations of material and methods for making synthetic dyes. Because one of the early and important materials used in these syntheses was aniline (derived from coal tar), the synthetic dyes have been known as coal-tar dyes and aniline dyes.

The manufacture of dyes proceeds from simple raw materials, mostly aromatic hydrocarbons such as benzene (benzol), toluene, and naphthalene, with introduction of other chemical groups, such as nitro, amino, halogen, and sulfonic acid. These intermediate compounds are then further processed by many special operations in organic chemistry such as diazotization, coupling, condensation, and fusion to give the final dyestuff. Following the chemical manufacture, the dye may be treated to give it special physical properties, and is standardized for strength. It may be sold as a dry powder, as a paste, or in solution.

Of the many thousands of dyes which have been synthesized in laboratories, about 3500 have had actual commercial use. These have been indexed in several compilations. The Society of Dyers and Colourists (England) and the American Association of Textile Chemists and Colorists give each dye an individual five-digit number as identification regardless of the specific manufacturer. Although some of the earliest dyes have common names with general usage, commercial dyes are sold under a large number of names, many of which relate to the same material.

Chemical structure. The most precise and scientific classification of dyes is based upon their chemical structure. This is the classification of interest to the research chemist and the manufacturer of dyes.

Utilization. This classification is based on the materials to which the dyes are applied.

Cloth. Coloring cotton, wool, linen, and silk is by far the largest use for dyes.

Manufactured fibers include regenerated cellulose (viscose), cellulose acetate, nylon, polyacrylonitrile, and polyester. Regenerated cellulose fiber can be dyed with dyes for cotton; dyeing other manufactured fibers with the colors commonly used for natural fibers is often less satisfactory or impossible. Special classes of dyestuffs were created for these polymeric materials. *See* TEXTILE.

Paper. This is colored both by dyeing and by pigmentation with finely divided colored materials. The color is usually added to the raw stock in the beater before the sheets are formed, or on the calender. Dyes are also applied in the coatings used on paper, such as the wax-dye coating used for duplicating (carbon) paper. *See* PAPER.

Leather. One of the earliest materials to be colored, leather has retained the use of natural dyes to a greater extent than most other materials, but there are many manufactured textile and specific dyes for leather. Because of the difficulty of penetrating the closely knit structure of leather, surface coloring by spraying or brushing is widely practiced. *See* LEATHER AND FUR PROCESSING.

Wood. Dyeing or tinting of wood is done with dyes in water, alcohol, or other solvents which evaporate after the wood has been painted with or soaked in the dye solution.

Pigments. Many soluble dyes are converted into pigments by forming insoluble salts (that is, replacing sodium in a dye salt with calcium) for use in lacquers, paints, and printing inks. This insoluble salt, called a toner, may also be deposited on inorganic fillers, such as aluminum hydroxide, to form a lake.

Food. The appearance of many foods is enhanced by artificial coloring. Butter and margarine are colored yellow, fruits and sauces which are often dulled by the process of preserving are brightened by addition of dye, and large amounts of dyestuffs are used in soft drinks and candies. In most countries, only certain colors are permitted to be used in this way.

In the United States, food must be colored by a natural dye or by a synthetic chosen from a list of certified food colors, of which the harmlessness to the human system is under constant check and review by a federal agency. Elaborate studies of the toxicity and pharmacology of any new color must be provided by the manufacturer before it is admitted to the list. *See* FOOD MANUFACTURING.

Oils. Lubricating oils and gasolines are colored for improving in appearance; for identifying grade or type of use; or as a trademark of the manufacturer. Waxes, shoe polishes, and candles are also colored.

Plastics and rubbers. Coloration of resins, plastics, and elastomers may be done by solution of a dyestuff or by dispersion of a pigment in the substrate. In this coloring operation, there is no need for substantivity, and choice of color is determined solely by the solubility, shade, and fastness properties desired.

Biological samples. The dyeing or staining of tissues of animal and vegetable matter is an important technique in physiological and medical studies. Microorganisms and cell structure are made more visible and differentiated under the microscope by selective staining with dyes of varying affinity for protoplasm.

Photography. Certain dyes (cyanines, of the polymethine class) are added to photographic emulsions to increase sensitivity to light in special regions of the spectrum. Color photographs are produced by formation of dyes from their components within the emulsion layer. *See* PHOTOGRAPHY.

Indicators. Some property of the environment may bring about a change in the hue in certain dyes, thus giving information to the observer about this property. Indicators usually respond to acidity, alkalinity, and oxidation or reduction, but may respond to heat, humidity, electricity, and water hardness. *See* ACID-BASE INDICATOR.

Miscellaneous. Soap, synthetic detergents, cosmetics, ink, hair, fur, metals, anodized aluminum, and many other materials are colored with dyes of various types. Dyes are also used to produce colored smokes, particularly for military identification. *See* INK; SMOKE.

Methods of application. This classification is used most frequently by the practical dyer.

Acid dyes. These are salts of organic acids (sulfonic and carboxylic) and are usually marketed as the sodium salts. The acid groups confer water solubility on the dyestuff molecule. When dissolved, the dye

ionizes (separates into particles with opposite electric charges), with the dye structure in the anionic (carrying the negative charge) part.

These dyes are used principally for wool, natural silk, synthetic fibers of polyamide and polyacrylic nature, leather, and paper. They are normally applied to the fiber in a solution containing some sulfuric or acetic acid, although some acid dyes will be fixed on the fiber from a neutral bath.

Because of improved fastness resulting from treatment of the dyed material with metal salts, especially chromium, a large number of acid dyes are available which contain metal atoms as part of the anion, as distinguished from the cationic metal atoms which form the salts of the organic acid. These anionically bound metal atoms do not exhibit the usual reactions of metal ions and are said to be chelated. These metalized (or premetalized) dyes have markedly superior fastness, approaching the vat dyes in this property. *See* CHELATION.

Some organic chemical compounds (mainly stilbene derivatives) have substantivity for fibers, but do not absorb light in the visible spectrum and hence show no color. They do transform some of the ultraviolet light which they absorb into visible light, thereby increasing the amount of white light reflected from them. This gives a bluing effect to yellowed materials and makes them appear whiter. Such products, used on cloth and paper, are called whites dyes, optical bleaches, or optical brighteners.

Basic dyes. These, too, are salts, but of organic bases containing amino and imino groups. The colored base is combined with a colorless acid, such as hydrochloric or sulfuric, and in solution, the dye structure is in the cation (carrying the positive charge). These dyes have exceptional brightness, but generally have only fair to poor fastness except when applied to acrylic fibers. They have wide utility for coloring wool, silk, leather, acrylic fibers, and paper.

Mordant dyes. Dyes which have little or no affinity for certain substrates may yet be fixed thereon if a mordant has first been applied. The fixation of the color is principally the result of reaction with the mordant material. These colors are also called adjective colors, as compared to the direct-dyeing substantive colors. Basic dyes which are applied to mordanted cotton are commonly still called basic colors, so that the designation mordant practically covers only acid-dye types and alizarin. The most common mordants are chromium salts for chrome dyeing processes. Often the chromium compound is applied during the dyeing, along with the dye, or it can be applied after the dyeing. These treatments may not only change the shade of the original dyeing (the self-shade), but they also may improve fastness of the dyeing to both water and light. *See* MORDANT.

Direct dyes. These dyes are normally sodium salts of sulfonic acids, and the colored part of the molecule is the anion. They differ from the acid dyes in that they are so substantive to cotton or other cellulosic fibers that they are fixed on the fiber from an aqueous solution with the assistance only of additions of common salt or sodium sulfate to the dyebath. Some authorities limit the designation substantive to these direct colors because of their outstanding affinity. These colors are also of importance in coloring paper, leather, and silk and have many miscellaneous uses. Because of the ease of application, they constitute the bulk of the package dyes used by the homemaker. This is an important class of dyestuffs, exceeded in numbers only by the acid dyes and in quantity used (in the United States) only by vat dyes.

Many direct dyeings can be improved in wetfastness by an aftertreatment such as with copper or chromium salts, formaldehyde, resins, and cationic fixing agents. It is possible to modify dyeings made by dyes of certain structures by a further chemical treatment (diazotization of a free amino group in the dye molecule, followed by coupling with a developer) of the dye on the fiber. Dyes suitable for such treatments are developed dyes. These treatments may or may not change the self-shade. *See* DIAZOTIZATION.

Ingrain dyes. These dyes are formed directly on the substrate by some type of chemical action. The principal subclasses are the azoic dyes and the oxidation dyes. In a few instances, dyes of the phthalocyanine type are developed on the fiber by special treatments.

Azoic dyes are water-insoluble azo compounds which have been formed within the substrate by chemical reaction of the intermediate components. Generally, the dyeing operation proceeds by dipping the cloth into a solution of one of the components (a hydroxy or amino component), drying the cloth without rinsing, and then treating it with a solution of the other component (usually a diazo component, which must be kept cold to prevent decomposition; hence, these colors are often called ice colors). Because the product of the reaction, which will then be deposited throughout the fiber, is a water-insoluble color, it is actually a pigment instead of a dyestuff. These colors are used for cotton and rayon, and in great quantity in printing.

Oxidation dyes are produced directly in the fiber or other substrates by a chemical oxidation following the impregnation of the substrate with certain aromatic amines. The final product is probably a pigment instead of a dyestuff. These colors are used principally for the dyeing of hair and fur.

Disperse dyes. These colors were originally developed for use on cellulose acetate, a synthetic fiber. This fiber has little affinity for the older known classes of dyestuffs. Some very slightly water-soluble colored materials transfer to this fiber from a water suspension if the colors are extremely finely divided particles. This results in a solution of the dye in the solid fiber.

Vat dyes. This class of colors is distinguished by the special method of application needed—a vatting operation, wherein the water-insoluble color is made soluble by a chemical reduction of the chromophore, the ketonic $=C=O$ group, to the $\equiv C-OH$ group, the leuco compound, which in the presence of alkali forms the water-soluble leuco salt, $\equiv C-OAlk$. This vat solution, which is often a different color from the original insoluble material, has affinity for

cotton, dyeing it with the shade of the vat solution. Upon oxidation with air or oxidizing agents, the reaction reverses to form the original water-insoluble color, leaving it deposited in the fiber as a pigment.

Sulfur dyes. Dyes of this class are also applied by a vatting technique which makes the solubilized color substantive to cellulosic fibers. These colors are made by treating a wide variety of organic compounds with sulfur and sodium sulfides. With a few exceptions, the final reaction products are not well-identified chemical compounds with known structures. These water-insoluble dyes are dissolved in alkaline sodium sulfide solution, which serves both as reducing agent and as source of alkali. After the dye is applied in the soluble leuco form, oxidation produces the insoluble dye on the fiber.

These colors are mostly used on cotton and viscose rayon. They have moderate all-around fastness and are relatively inexpensive. Hence, they are used in large quantities.

Solvent dyes. These colors are soluble in organic solvents such as benzene, gasoline, alcohol, acetone, oils, fats, and waxes. Solvent dyes color merely by solution of the dye in the substrate. They may be subclassified as spirit-soluble colors with principal solubility in alcohol and as oil-soluble colors which are soluble in benzene and vegetable and mineral oils. Uses include wood stains and varnishes, lacquers, printing and writing inks, butter and margarine, and plastics and resins.

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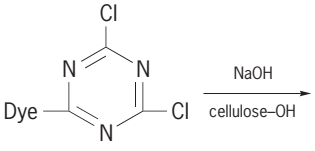
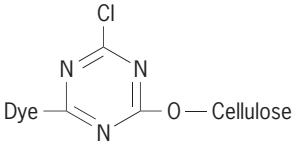
Reactive dyes. Dyes that during the process of dyeing form a covalent bond with the substrate are known as reactive dyes. They are of particular interest on cellulosic fibers, where good washfastness is obtained without resorting to the production of large insoluble agglomerates within the fiber such as occur with vat and azoic dyes.

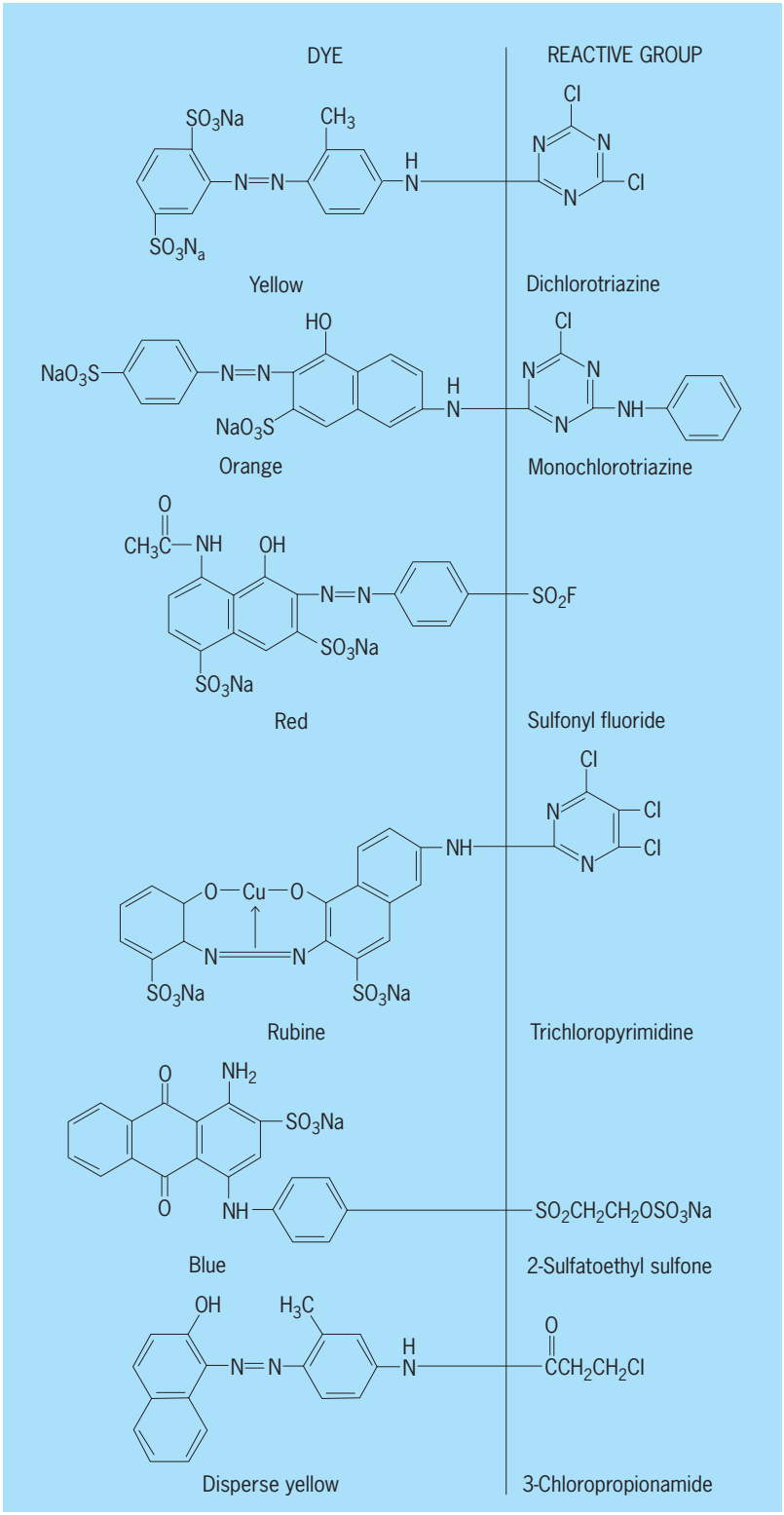
Structure. A reactive dye (see **illus.**) is a soluble color that contains a group capable of reacting with the hydroxyl or amino groups in cellulose, wool, silk, and other substrates. The reaction must be rapid and be carried out in the presence of water. Generally, an acid acceptor and heat are used to promote the reaction. Reaction efficiency varies from 60 to 95%, and unreacted dye must be removed by scouring to ensure washfastness.

Two types of reactive groups have been found successful in dyes which combine chemically with cellulose: active-halogen compounds which form ester or amide links [reaction (1)], and





(1)



Structural formulas for typical reactive dyes.

activated vinyl groups which form ether links [reaction (2)].

$$\text{Dye}-\text{SO}_2\text{CH}_2\text{CH}_2\text{OSO}_2\text{Na} \xrightarrow{\text{NaOH}} \text{Dye}-\text{SO}_2\text{CH}=\text{CH}_2 \xrightarrow{\text{cellulose-OH}} \text{Dye}-\text{SO}_2\text{CH}_2\text{CH}_2\text{O}-\text{Cellulose} \qquad (2)$$

Application. Reactive dyes can be applied by either continuous or batch techniques. The most popular continuous methods are the pad, dry chemical pad, steam and scour, or pad thermofix and scour procedures. Batch application involves exhausting the dye with salt and then introducing the acid acceptor to promote reaction with the fiber. The method of application depends on the type of reactive groups and on the material being dyed. Nitrogenous fibers generally react at neutral or slightly acid conditions, and hydroxy-containing fibers require an alkali to promote reaction. When the reactive group is stable in the presence of the acid binder at room temperature, the process, becomes simply pad, heat, and scour.

Use and potential. The major penetration of reactive dyes has been into the cellulosic fiber dye market, as they offer improved washfastness over direct dyes, improved brightness over vat dyes, and improved rubbing fastness and shade range over azoic dyes. Reactive dyes are one of the more important single types of cellulosic dye. To a lesser degree, they are used for polyamide fibers.

The simplicity of continuous application and the brightness of reactive dyes make them useful in printing as well as dyeing. Both prints and dyed fabrics may be resin treated without significantly affecting the dye properties when selected resins are used. *See* HETEROCYCLIC COMPOUNDS; QUINONE.

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Dyeing

The application of color-producing agents to material, usually fibrous or film, in order to impart a degree of color permanence demanded by the projected end use. True dyeing covers mechanisms in which molecules of material to be dyed become involved by various means with the molecules of the coloring matter, or small aggregates thereof. There is some overlapping between true dyeing and other methods of coloring, which are called dyeing in the industry. Products which are commonly dyed include textile fibers, plastic films, anodized aluminum, fur, wood, paper, leather, and some foodstuffs.

The broad term affinity is used to describe the various types of attraction between the material to be colored and the dye. Affinity may be caused by attraction between charged dye particles and oppositely charged dye sites on the material, by various types of chemical attraction, and by formation of solid solutions of dye in the material.

When affinity is involved, dyeing is an exothermic process which may be simply stated as follows: dye in solution + undyed material → dyed material + heat.

This simplified equation explains the universally known fact that dye has a greater tendency to bleed into hot water than into cold water. It also points out that more dye will ultimately be absorbed by a fiber or film at low temperatures than at higher temperatures when equilibrium is finally reached. However, the rate of dyeing increases geometrically, whereas maximum dye absorption decreases only arithmetically with rise in temperature. Maximum dye absorption is rarely necessary or desirable; therefore, the trend in modern practice is toward dyeing at high temperatures and short times.

Dyeing is accomplished by dissolving or dispersing the colorant in a suitable vehicle (usually water) and bringing this system into contact with the material to be dyed. Although many dye molecules or aggregates may adhere to the material surface when they meet, dyeing does not occur until the adhering dye particles migrate within the fibers of films. All dyeing processes are designed to accomplish ultimately penetration of the undyed substance by the colorant.

Assistants. These are materials which do not impart color to the product to be dyed but, as the name suggests, promote or retard dyeing. Usually, but not always, they affect the dye molecule.

Swelling agents are assistants which open up the structure of the fiber temporarily so that dye molecules or aggregates may enter more freely and reach otherwise inaccessible dye sites.

Carriers are agents (often solvents of low water solubility) which accelerate dyeing by breaking up or dissolving dye aggregates and bringing them to the fiber-water interface in a size small enough to be absorbed by the material. Frequently, a carrier may exert a swelling action on the fiber in addition to its normal function.

Dye retarders are a class of dyeing assistants, usually inorganic or organic salts, which slow up the dyeing process by forming evanescent compounds with the dye, by buffering or depressing the ionization of an acid assistant, or by temporarily occupying the more active or more accessible dye sites on the fiber, later to be dislodged therefrom by the dye.

Aftertreating agents are salts, resins, or other products (more frequently applied to cellulosic fibers) to render the colored fabric more resistant to the effects of washing, perspiration, or fading by ozone or combustion gases. More often than not, their application causes a loss in lightfastness of the dyed material. Aftertreatment with copper salts, for example, is normally in order to increase lightfastness, whereas application of formaldehyde-urea or melamine resins to increase wetfastness normally affects sunlight resistance adversely.

Cellulose fibers. Cotton and rayon are most commonly dyed by immersion of the fibers in a solution of direct dyes using an electrolyte such as common salt as assistant and then boiling this dyebath. The affinity in this case may be a result of hydrogen-bonding of areas in the dye molecule to hydroxyl groups in the cellulose, more simply stated as adsorption. Such dyeings usually exhibit only commercial

(minimum) resistance to washing. Treatment of the properly dyed fibers with resins and copper, for example, increases the resistance to washing with minimum loss of light resistance. Some of the direct dyes have chemical groups (amino) which enable them to be converted after dyeing to large, more insoluble molecules by treatment with nitrous acid (diazotization) and a phenol or an amine (developing). Such dyeings are quite resistant to degradation by washing, but are usually characterized by poor resistance to fading by sunlight. Naphthol dyeing is a special form of developed dyeing wherein the naphthol is a selected type of substantive developer and is applied to the fiber first. Then a so-called diazo salt (not a dye) is added and an insoluble pigment forms within the fibers. *See* NATURAL FIBER.

Cellulose fibers may be dyed by vat dyes. These dyes are normally insoluble pigments. However, under the influence of alkali and reducing agents, they become water soluble and exhibit affinity for various fibers. After dyeing, the vat dyes revert to their insoluble form upon exposure to air or to oxygen-supplying chemicals. Indigo is a vat dye which does not possess affinity. Its alkaline-reduced solution, therefore, can be impregnated into the fiber only by successive dips and oxidations. Most vat dyes when properly applied are extremely resistant to degradation by such agents as washing, bleaching, sunlight, and perspiration.

Cellulose may be dyed by so-called fiber-reactive dyes which form tenacious bonds to cellulose and actually convert the cellulose to a different compound. The connecting links between dye and fiber are considered to be ether or ester linkages.

Conventional sulfur dyes are applied to cellulose from a sodium sulfide solution and are characterized by subdued hues, good washfastness, and low bleach resistance. They are widely employed for the production of deep shades on cottons destined for use in work and play clothes. Sulfur dyes made soluble by sodium thiosulfate are now available. These dyes are applied continuously to cellulose in the oxidized state (no affinity at this point), then reduced in a steamer thereby regaining "affinity" and thus dyeing the goods. Some of the newer "thiosulfate" or "condense" dyes produce hues far brighter than those obtainable with conventional sulfur dyes.

Basic dyes can be dyed upon cellulose which has previously been mordanted with synthetic tannins or tannic acid and tartar emetic. Dyeing is accomplished by the formation of a "lake," or pigment of the dye and mordant; the function of the cellulose is merely that of a substrate in this type of dyeing. Such dyeings are brilliant in hue but low in light resistance, and, for the latter reason, the use of basic dyes on cotton has become obsolete.

Animal fibers. The dyeing of wool, silk, and fur (felt) involves the formation of salts by reactive positively charged groups on the fiber and negatively charged color groups on anionic dyes.

Level-dyeing (acid or anionic) dyes, which require a relatively large amount of strong acid to force the dye onto the fiber, are used for carpet yarns, felt hats, and wherever even dyeing and penetration instead of washing fastness is a paramount requirement. Milling dyes, which demand less acid, are faster to washing and are used for blankets and sweaters. Where a very high resistance to light and washing is required, acid-dyeing or neutral-dyeing metallized dyes are employed.

Chrome (or mordant) dyes are applied to wool where subdued hues of excellent fastness are needed. In the metachrome method, chrome (sodium bichromate) is added to the dyebath at the start of the dyeing process. If the wool is treated with the chrome mordant before dyeing, the term chrome-bottom is used; if the chrome is added at the end of the dyeing cycle, the term top-chrome or after-chrome dyeing is applied. If silk-effect threads (such as in pin-striped cloth) are present, the afterchroming is performed in a separate fresh bath, a technique called the silk-white process.

Vat dyes and vat esters are also dyed on wool when bright shades of extreme fastness are required. The leuco vat esters are developed by acid and oxidizing agents, and the vat dyes are dyed as on cellulose but at lower temperature, lower alkalinity, and with a greater amount of reducing agent (sodium hydrosulfite). Fiber-reactive dyes for wool have been developed.

Manufactured fibers. Cellulose acetates and triacetate (Arnel) are dyed in a supension of solvent-soluble dyes by immersion (disperse-acetate). These dyes are considered to form solid solutions in cellulose acetate by passing from the water phase to the water-fiber interface and migrating from the fiber surface inward. In a technique similar to the diazotizing and developing technique described under cellulose dyeing, water-insoluble dye precursors are applied to the cellulose acetate; diazotization is accomplished with sodium nitrite and acid; β-hydroxynaphthoic acid or hydroxanil are added to develop an insoluble colorant within the fiber. Cellulose acetate may also be dyed by immersion in alcohol-water solutions or in formic acid-water solutions of certain water-soluble acid dyes.

Polyamide fibers are dyed like wool with acid, metallized acid, neutral metallized, fiber-reactive mordant dyes, azoics, and selected direct dyes from an acid bath. Vat dyes are applied by methods used for cotton dyeing, but higher temperature and swelling agents such as *o*-phenylphenol are used. Disperse dyes are dyed on nylons when maximum freedom from streaks (hosiery and flat woven goods, for example) is a prerequisite and resistance to washing and light is less important.

Acrylic fibers such as Orlon and Acrilan-16 are dyed in light shades with disperse (acetate) dyes at high temperatures. Being negatively charged, the acrylics are most successfully dyed to heavy hues with cationic (positively charged) colorants. By simultaneous use of copper salts and reducing salts, acrylics may be dyed with anionic dyes (the so-called cuprous-ion method). The positive cuprous ion forms a link between the negative dye molecule and the negative fiber molecule.

Modified acrylics (Acrilan, Zefran, Creslan, and Verel, for example) are dyed with disperse and basic dyes, as well as with dyes normally applied to natural fibers. The modification of the fibers involves the introduction of basic materials which form dye sites for anionic dyes.

Polyester fibers (Dacron, Terylene, Kodel, and Fortrel) are dyed with disperse dyes at high temperatures or by the use of swelling agents and carriers such as salicylic acid or chlorinated solvents. These fibers may also be dyed under pressure at high temperatures by leuco vat esters and disperse dyes or by applying these same dyes, drying, and passing through a hot flue or over heated rollers at about 400°F (204°C; Thermosol method).

Polypropylene fibers have been dyed in medium shades with leuco vat esters and sulfur dyes from aqueous baths at high temperatures by using thiourea dioxide as a reducing agent and diammonium phosphate and dispersing agents as dye assistants. Azoics from an acid bath and vat acids are also applicable. Solvent vapors have been used to dye polypropylenes with solvent soluble, oil, disperse, and related dyestuffs.

Polypropylene has been altered by grafting or with copolymers which permit dyeing the modified fiber with anionic or disperse dyes. A unique coloring method involves the use of dyes which can chelate onto nickel. (Some polypropylene brands contain nickel stabilizers.) Such chelate dyeings and prints have extremely high sunlight fastness. *See* MANUFACTURED FIBER.

Pigment dyeing. All fibers may be colored by application of pigments dispersed in a bonding agent. Affinity or true dyeing is not involved. Good fastness of pale shades to most normal color-destroying effects and of heavy shades to all but rubbing-off is characteristic; however, the bonding agent may stiffen sheer goods.

Dope or spin dyeing. Melt spun fibers such as polypropylene, nylon, Dynel, or Dacron may be colored in the mass. The colored plastic is melted and then extruded through spinnerets (refined shower heads), and the congealed streams are collected and processed to yarn quality.

Fibers such as cellulose acetate, acrylics, and viscose rayon are prepared by extruding their solutions into hot air or a coagulating water bath. These so-called dopes are colored with pigments before extrusion, and the coagulated fibers collected and processed into yarn.

Textile-dyeing equipment. There are two fundamental types of dyeing processes—the dye liquor is pumped through stationary material, or the material to be dyed moves through the dye liquor.

Kiers are tanks which hold loose fiber or packaged yarn and the dye liquor is pumped through the fiber at temperatures up to the boiling point or (if the fiber can withstand it) under pressure at temperatures up to 270°F (132°C). In one machine, goods are wound upon a perforated roller (or beam), and the whole placed in a pressure chamber. Dye liquor under pressure above 212°F (100°C) is pumped through the goods. Several other machines which function by pumping dye liquor through stationary material are used in modern mills.

In continuous dyeing, cloth is impregnated with dye and then passed through a series of developing, washing, and drying zones to a final take-up roll. One machine uses a passage through molten Wood's metal to develop the dye-saturated goods, whereas another uses hot oil for this purpose.

The pad-stream system consists of rollers which apply dye and necessary chemicals to the goods; development takes place on continuous passage through a steam chamber.

The pad-roll system uses an insulated oven in which a huge roll of goods, previously saturated with dye solution, slowly turns until the dye has become fixed to the fiber.

A jig promotes dyeing by winding goods through the dye bath, back and forth from one roll to another.

A beck consists of an elliptical reel which draws goods which have been sewn in an endless chain from the dye bath and plaits it back into the bath repeatedly until dyeing is complete. The use of a machine equipped with disked roller-pressure locks permits continuous dyeing at high pressures.

A relatively new machine applies dye continuously to carpet faces by means of a roller which picks up dye solution and then delivers this film of solution to be scraped off by a fluorocarbon "doctor." The resultant cascade of dye solution falls onto the carpet face, and the carpet moves forward into a steamer. Gum thickeners and a foam generator (possibly ethyl carbonate) hold the dye solution evenly throughout the pile height enabling even and nonfrosty (light pile tips) dyeing. A somewhat similar machine employing a padder and steamer is used for tow or a loose web of fibers. The latter's practical use predates the carpet dyeing machine. It employs the principle of coacervation to effect rapid dye fixation. Equipment which continuously develops disperse dyes on polyesters by dry heat (thermofixation) is wisely used. Dyeing continuously in solvent solutions or in solvent vapors is being investigated to speed up production and to minimize pollution.

Nontextile materials. Dyeing anodized aluminum was formerly believed to be accomplished by the formation of aluminum salts of dyes by the electrically deposited aluminum oxide coating. However, many dyes having no salt-forming groups were found to dye this film, and it is postulated that, at least in these cases, conical craters receive the colorant, and subsequent sealing by boiling in water or in solutions of salts in water converts the oxide to larger hydrated molecules (for example) and imprisons the dye aggregates in the sealed-off cones. At any rate, anodized aluminum is readily dyed by many textile dyes. Light and weather resistance undreamed of in textile applications of some of these same dyes is achieved.

Paper. Paper pulp is usually dyed in the paper beater by dyes normally employed for cotton; on occasion, it is tinted by wool dyes, and it is frequently tinted by addition of pigments to the beater. Finished

paper is also colored by passing it over rollers which supply dye or colored coatings to its surface (calender staining). *See* PAPER.

Wood. Wood is normally stained with solutions of dyes or dispersions of pigments in water, in solvents, or in lacquers. Wood, however, is frequently dyed by the application of water solutions of dyes at high temperature under pressures of 100 lb/in.^{2} (1700 kilopascals) or more. Freshly felled trees have been dyed by force-pumping dye solution through the length of the log. The dye solution replaces the natural moisture and sap. An inflated tubular gasket seals the log-end to the pressure cylinder in this method. Wood, destined to be cemented to corks for liquor bottle closures, has been dyed with sulfur dyes to ensure resistance to alcohol.

Monomeric substances, such as methyl methacrylate or styrene, are colored with solvent-soluble dyes. These colored solutions are impregnated into wood, and the corresponding polymer is formed therein by radiation catalysis. Such wood is extremely durable and marproof since every fibril throughout the wood is "varnished" by this process.

Leather. Leather is dyed at low temperatures with the classes of dyes normally used for wool and cotton, excepting vats and naphthols. Formic acid is normally used to exhaust the dye. For dress gloves, leather is usually colored by applying the dye on the grain surface, leaving the flesh side undyed. Leather is also dyed with natural dyes such as logwood, fustic, and quercitron. Leather fresh from tanning and containing considerable moisture is dyed in Europe by tumbling with dry water-soluble dye. Greater penetration by the dye is claimed by this process as opposed to normal dyeing in water solutions of the dyes. *See* LEATHER AND FUR PROCESSING.

Food products. Most food products which are artificially colored are not actually dyed. Maraschino cherries, however, are dyed for several hours with food dyes, then washed and placed in flavored syrup. Glazed citron, citrus, or watermelon rinds are colored by long immersion followed by drainings in a series of food-dye-colored, increasingly strong, sugar syrups. Pistachio nuts in the shell are dyed by applying food-dye solutions with or without salt during roasting operations.

Fur. Fur is usually colored by impregnations with synthetic organic intermediate compounds (amines) which are then oxidized to colored pigments (azines) with peroxides, for example. When not fully developed, some of the products produce objectionable physiological reactions so that great care is taken to assure the absence of unreacted amine in the dyed fur.

Human hair. It is not generally realized that human hair is, in the final analysis, a class of textile fiber. Cosmetic companies, appreciating this fact, have hired textile experts to do research on dyeing, waving, and washing hair. The dyeing of hair is of course limited to processes employing relatively low temperatures. Food dyes (also carbon black) mixed with citric or malic acid are used for so-called rinses. Other rinses employing basic dyes and salt or textile dyes diluted with salts, detergents, and malic, citric, or tartaric acids are on the market. Products selected, as a rule, from among the safer fur-dye intermediates are used for more permanent and intense coloring techniques; these employ hydrogen peroxide for developing. Natural dyes, such as the age-old henna and some metallic salts, are still encountered in some hair dyeing establishments. Rinses which are used to mask the yellowish tinge in gray or white hair frequently use blue or violet food or textile dyes and contain optical bleaches—substantially colorless compounds which receive invisible ultraviolet light and reflect it to the eye as a blue or violet fluorescence.

Plastic films or products. Many plastic materials may be dyed by processes similar to those for textiles. Nylon, cellulose acetate, polyethylene, polypropylene, and polyester resins are dyeable with dyes which color these materials in yarn form. Solvents such as pyridine, dioxane, alcohol, and trichlorobenzene, when applied in solution or emulsion form, swell these and other resins as an aid in applying solvent-soluble dyes. Research is under way to develop and improve reactive dyes for polycarbonate and other new resins. *See* DYE; PLASTICS PROCESSING; TEXTILE CHEMISTRY.

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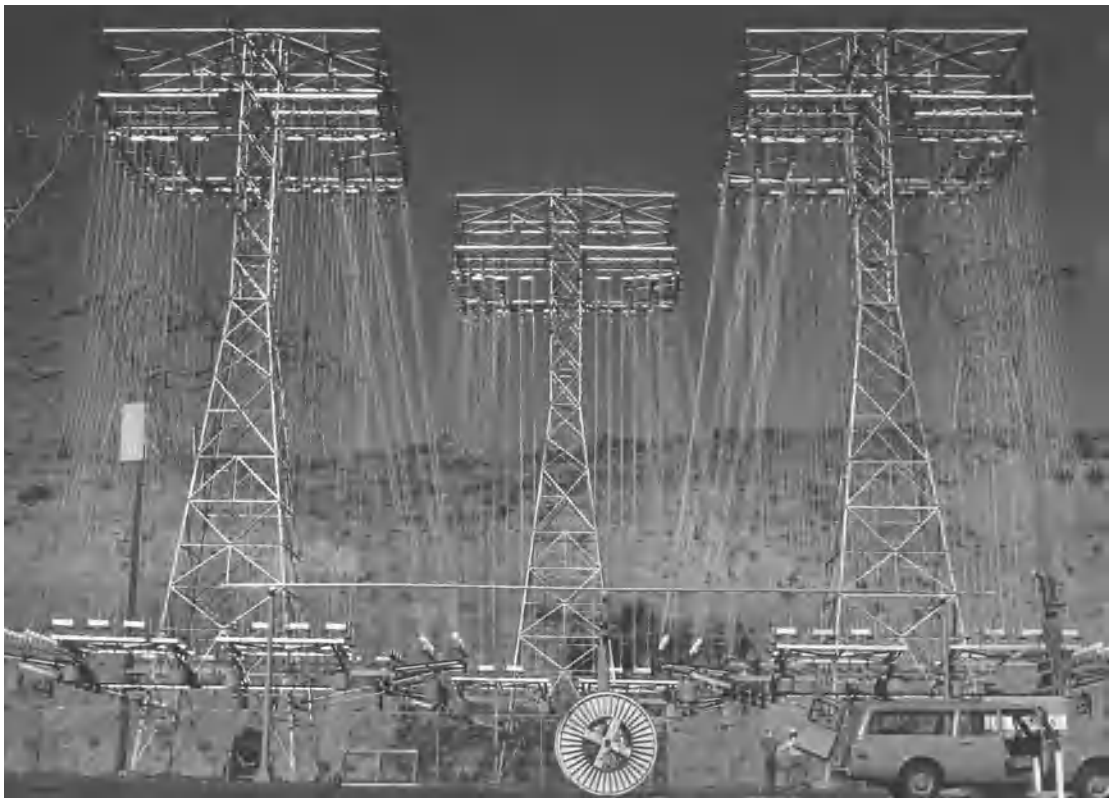
Dynamic braking

A technique for braking in which mechanical energy is converted to heat or electrical energy in order to slow or stop motion. An all-mechanical dynamic brake consists of rotating vanes that circulate a viscous fluid in a manner that generates heat. This is one way that the power of the wind is harnessed for space heating. An electric dynamic brake consists of an electric dynamo in which the mechanical energy is converted to electric form, and either converted to heat in a resistor or returned to the supply lines. Typically, electric braking is accomplished with the same machine that serves as the drive motor. Electric dynamic braking is employed in electric vehicles, elevators, and other electrically driven devices that start and stop frequently. *See* WIND POWER.

The most common type of dynamic braking will be explained for a direct-current (dc) motor. To accomplish braking action, the supply voltage is removed from the armature of the motor but not from the field. The armature is then connected across a resistor. The electromotive force generated by the machine, now acting as a generator driven by the mechanical system, forces current in the reverse direction through the armature. Thus a torque is produced to oppose



Dynamic braking resistor with a capacity of 1400 MW, installed in the Bonneville Power Administration's Chief Joseph Substation. (*From M. L. Shelton et al., Bonneville Power Administration 1400-MW braking resistor, IEEE Trans. Power Apparatus Sys., PAS-94:602–611, Institute of Electrical and Electronics Engineers,1975*)

rotation, and the load decelerates as its energy is dissipated, mostly in the external resistor, but to some extent in core and copper losses of the machine. *See* DIRECT-CURRENT MOTOR.

Electric braking can also be accomplished by causing the energy of the rotating system to be converted in the armature to electrical energy and then returned to the supply lines. This mode of operation, called regenerative braking, occurs when the counterelectromotive force exceeds the supply voltage.

Interchanging two of the lines supplying a three-phase alternating-current (ac) induction motor also produces braking. In this case, called plugging, the direction of the electromagnetic torque on the rotor is reversed to cause deceleration. Both the energy of the system and the energy drawn from the supply lines are expended in copper and core losses in the machine as heat. The power lines must be disconnected when the rotor comes to rest. *See* INDUCTION MOTOR.

Regenerative braking is a system of dynamic braking in which the electric drive motors are used as generators and return the kinetic energy of the motor armature and load to the electric supply system. This method is employed when the load is losing a large amount of potential energy, as in the case of an electric train descending a long grade or a pumped storage reservoir being emptied. The potential energy of the load is first converted to kinetic energy in the motor armature and load and then to electrical energy. The method is applicable to dc motors, to induction motors (with negative slip), and to synchronous motors. *See* PUMPED STORAGE; SYNCHRONOUS MOTOR.

Rheostatic braking is a system of dynamic braking in which direct-current drive motors are used as generators and convert the kinetic energy of the motor rotor and connected load to electrical energy, which in turn is dissipated as heat in a braking rheostat connected to the armature. This is accomplished by disconnecting the motor armature from the supply lines and connecting them to a braking rheostat, while the field circuit remains connected to the supply line. The polarity of the armature remains unchanged, but the direction of the armature current reverses, resulting in a negative torque. The torque can be controlled by means of the braking rheostat.

Sometimes the term dynamic braking is applied only when mechanical energy is dissipated in an external resistor, but its more general interpretation includes regenerative braking and plugging.

Dynamic braking is also applied in electrical utilities as an aid in stablizing operation of interconnected systems. A fault on an electric line has the effect of reducing both the electric power out of a generator and the power transmitted through an interconnection tie line. The prime mover governors, particularly for hydroelectric generators, cannot respond fast enough to avoid acceleration of the speed of one interconnected generator relative to another. The machines no longer operate synchronously, and separation of the interconnection points results.

Dynamic braking resistors can be rapidly connected to a generator bus as a dummy load until the fault is cleared and the normal power flow through the interconnection is reestablished. In this way the acceleration of the generator may be limited so that the synchronizing power of the tie line can maintain stability. The resistors employed for this purpose are huge. The dynamic braking resistor installed in Bonneville Power Administration's Chief Joseph Substation has a capacity of 1400 MW (see **illus.**). *See* ALTERNATING-CURRENT GENERATOR; ELECTRIC POWER GENERATION; ELECTRIC ROTATING MACHINERY; TRANSMISSION LINES. Arthur R. Eckels

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Dynamic instability

A state of fluid flow in which the distribution of mass and momentum is unstable. Fluid flows are subject to a variety of instabilities which generally cause the flow to become more complex and which often lead to turbulent, chaotic flow. Instabilities are responsible for a variety of phenomena in natural flows, including cyclones, hurricanes, and thunderstorms in the atmosphere; mantle convection in the Earth's interior; and granules and supergranules in stellar atmospheres. A great deal of research in the geophysical and astrophysical sciences has focused on flow instabilities; and instability plays an important role in engineering problems ranging from naval engineering and aeronautics to the design of efficient heating and cooling systems. *See* FLUID FLOW; TURBULENT FLOW.

Fluid instabilities may be broadly divided into two classes: convective and dynamic. While convective instabilities can be understood rather easily in terms of forces acting on displaced parcels of fluid, dynamic instabilities are more varied and more challenging to understand. This broad class of fluid instabilities is responsible for the cyclones and anticyclones that dominate the weather in middle and high latitudes, as well as meanders and rings in ocean currents such as the Gulf Stream. These instabilities are ultimately responsible for the lack of predictability of complex flows such as are found in the atmosphere. *See* CONVECTIVE INSTABILITY; CYCLONE; WIND.

Vorticity and invertibility. Many, but not all, dynamic instabilities can be understood with the aid of a vorticity principle and an invertibility principle. The vorticity of a fluid can be thought of as the rate of rotation of a rigid paddle wheel embedded in the flow, as illustrated in **Fig. 1**. (Mathematically, it is the curl of the vector velocity field.) The rotation can be brought about by shear (a change in the speed of the flow in the direction across the flow) and by the curvature of the flow. Vorticity is important for two reasons. First, in two-dimensional flows, vorticity is conserved; that is, if the paddle wheel is followed around it is observed that its rate of rotation remains

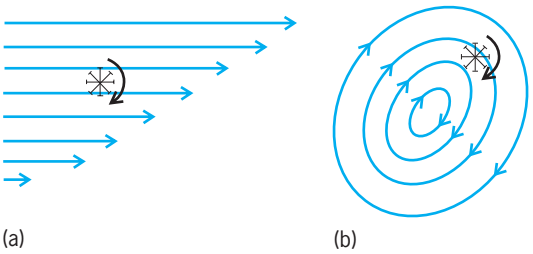

Fig. 1. The vorticity of a fluid is related to the rate at which a rigid paddle wheel would rotate if placed in the flow. (*a*) The paddle wheel rotates because the velocity of the fluid varies in a direction perpendicular to the flow. (b) Rotation is generated by the curvature of the flow.

constant. Second, in such flows, it is possible to work backward from knowledge of the vorticity at every point in the fluid to find the flow field at every point. This is called the invertibility principle. Basically, a clockwise vorticity at a point in the fluid will induce a clockwise rotation of the fluid nearby; the total velocity field in the fluid can then be found by adding the velocities induced by all the vorticities at each point. *See* CALCULUS OF VECTORS.

Barotropic instability. A shear flow is shown in **Fig. 2***a*. All the shear in this case is concentrated on a line, and thus all the vorticity is concentrated on the same line. Therefore, this flow can be perceived as being composed of a line of vortices (Fig. 2*b*). However, this flow turns out to be unstable, in the sense. If two of the vortices are displaced in the manner shown in Fig. 2*b*, the vorticity is conserved so that the rate of rotation of the vortices does not change. However, the flow induced by each of the two displaced vortices moves the other vortex in the same direction as it was displaced in originally. Thus the slightest perturbation of the original line of vortices will amplify rapidly, changing the whole flow. This is called barotropic instability, and is responsible for such phenomena as meanders in the Gulf Stream. Such instabilities draw on the kinetic energy of the original shear flow. A necessary condition for barotropic instability is that the vorticity has a maximum or minimum within the flow. Generally, an extremum strung out in a line such as in the example in Fig. 2 is unstable, while circular patches of high or low vorticity can often by stable. *See* BAROTROPIC FIELD.

Baroclinic instability. The flow of the Earth's atmosphere is three-dimensional. Under this

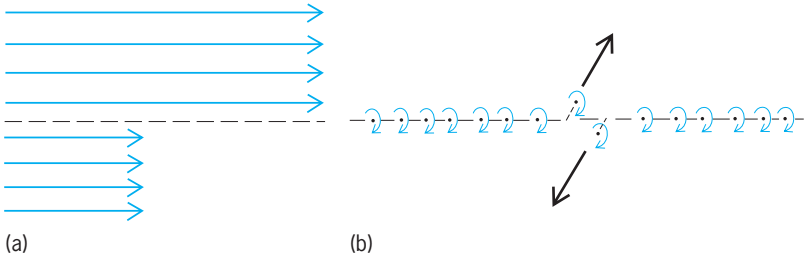

Fig. 2. Diagram showing barotropic instability. (*a*) All of the vorticity of an idealized shear flow shown concentrated on a line. (*b*) Line envisaged as a row of individual vortices. If two vortices are displaced in the manner illustrated in *b*, the flow they "induce" at each other's position enhances the original displacement.

circumstance, vorticity is not conserved because fluid converging on or diverging from a point in the fluid causes it to spin faster or slower. (The principle at work here is the conservation of angular momentum.) Fortunately, the Earth's atmosphere is usually stably stratified; that is, a parcel displaced upward will be colder than its environment and will accelerate downward toward its original position. In this sense the atmosphere is convectively stable. It is possible to identify a quantity that reflects the air's density and at the same time is conserved as long as no heat is added to or subtracted from the air. This quantity is called the potential temperature (θ) and is related to temperature and pressure by the expression

$$\theta = T\left(\frac{1000}{p}\right)^{0.287}$$

where T is the temperature in kelvins and p is the pressure in millibars. Generally, θ increases upward in the atmosphere: slowly in the troposphere (a layer of air that extends upward to roughly 6 mi or 10 km) and more rapidly in the stratosphere (a layer of air between about 6 and 18 mi, or 10 and 30 km). *See* STRATOSPHERE; TROPOSPHERE.

Because θ is conserved, a layer of air between two different surfaces of constant θ (which are nearly horizontal surfaces in the atmosphere) is always trapped between those surfaces. (If air were to cross θ surfaces, this would imply that θ is changing as the air is followed around.) If it is observed that two θ surfaces are moving farther apart, the implication is that air must be converging horizontally to fill the gap. If a floating paddle wheel could be placed on a θ surface in this circumstance, it would be observed that its spin increases with time, because of the convergence of the flow. The farther apart the surfaces are, the greater the rate of spin, or vorticity. In fact, the vorticity is directly proportional to the distance between two θ surfaces. This can be used to define a conserved quantity called potential vorticity, which is simply the ratio of the vorticity to the distance between θ surfaces. (Mathematically, it is the dot product of the curl of the vector velocity field with the gradient of θ, all divided by density.) An interesting aspect of potential vorticity is that for sufficiently large-scale flows, such as the low- and high-pressure systems that affect daily weather in middle and high latitudes, it also has an invertibility principle that is quite analogous to the invertibility principle for vorticity in two-dimensional flows. In the case of three-dimensional flows, a blob of high-potential vorticity will induce a counterclockwise circulation of fluid around it and above and below it (in the Northern Hemisphere).

The criterion for the dynamic instability of large-scale three-dimensional flows is also analogous to barotropic instability of two-dimensional flows. The necessary condition for this kind of instability is that potential vorticity must be a maximum or minimum somewhere on a surface of constant θ; once again, extrema of potential vorticity arranged along a line are almost always unstable, whereas circular blobs are often stable. The type of instability that results from maxima or minima of potential vorticity in a fluid is called internal baroclinic instability. This type of instability can draw on both the kinetic energy of the original flow and the potential energy that occurs when there are horizontal variations of density.

Observations of the Earth's atmosphere show that the potential vorticity of the troposphere in middle and high latitudes is nearly uniform. But when a gently sloping θ surface is followed into the stratosphere, a large increase of potential vorticity occurs. Even so, there is no localized maximum or minimum of potential vorticity, but just a large gradient of the quantity, concentrated near the boundary between the troposphere and stratosphere. (Many meteorologists define the stratosphere as a region of large potential vorticity.) For this reason, internal baroclinic instability seldom occurs in the atmosphere.

In order to understand the dynamic instability that causes cyclones and anticyclones (the familiar low- and high-pressure weather systems on weather maps), it is necessary to introduce a corollary to the invertibility principle, which recognizes an important and unique role of solid boundaries (such as the Earth's surface) within a fluid. This corollary states that the temperature of the fluid at the boundary behaves just as the potential vorticity does away from boundaries. That is, a patch of high temperature at the Earth's surface induces a counterclockwise rotation of the air around it, just as though this air were a blob of high-potential-vorticity fluid. Because of this corollary, it turns out that a condition for dynamic instability in the Earth's atmosphere is that the temperature gradient at the Earth's surface have the opposite sign of the gradient of potential vorticity on θ surfaces. Since surface temperature generally decreases poleward, while potential vorticity on θ surfaces increases poleward as the gently sloping θ surfaces are followed into the stratosphere, this condition is usually met in middle and high latitudes. The resulting instability is referred to simply as baroclinic instability.

The mechanism of baroclinic instability can be understood with the aid of **Fig. 3**, which shows the formation of a surface low-pressure system in the Northern Hemisphere. First, a tongue of high-potential-vorticity air downward and southward from the stratosphere. If this tongue is big enough, the counterclockwise circulation it induces extends all the way to the surface, where the tongue advects warmer air northward along the ground to its east. Once a tongue of warm air is formed at the surface, it also induces a counterclockwise circulation around it (by the corollary to the invertibility principle) which extends upward to the stratosphere. This counterclockwise circulation pulls down more high-potential-vorticity air from the stratosphere, reinforcing the original tongue, and so on. This positive feedback mechanism results in an amplification of the whole system; this is baroclinic instability. Also, the surface low-pressure center (the site of maximum counterclockwise circulation) is just west of

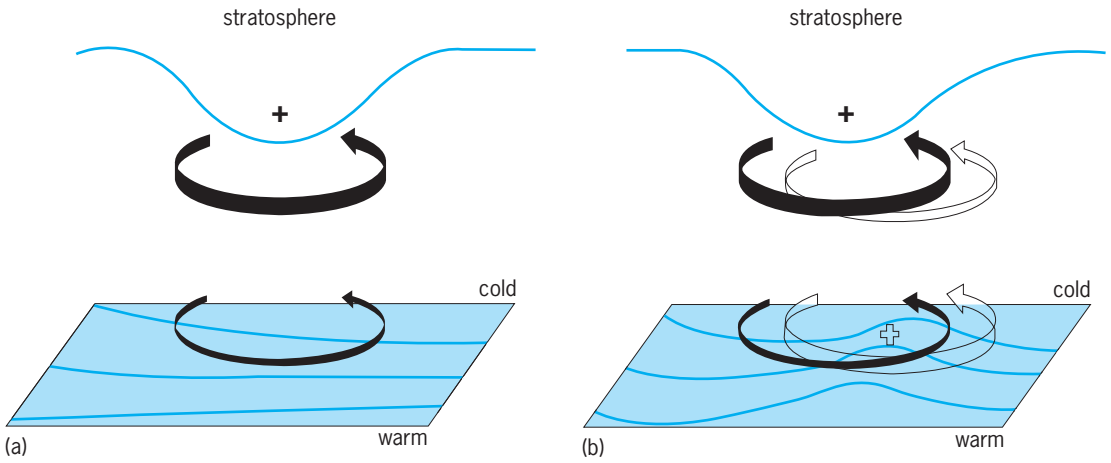

Fig. 3. Schematic diagram showing the formation of a cyclone in the Northern Hemisphere. (*a*) A tongue of high-potential-vorticity air from the stratosphere extends downward and southward into the troposphere; this tongue induces a counterclockwise circulation at the surface, where temperature decreases northward. (*b*) The induced circulation advects warm air northward, and the warm tongue induces a counterclockwise circulation of its own. This acts to pull more stratospheric air southward and downward, reinforcing the original perturbation. (*After P. J. Hoskins, M. E. McIntyre, and A. W. Robertson, On the use and significance of isentropic potential vorticity maps, Quart. J. Roy. Meteorol. Soc., 111:877–946, 1985*)

the tongue of warm air at the surface, which is where the circulations due to the warm-surface tongue and the high-potential-vorticity tongue aloft reinforce to the maximum extent. High-pressure systems form analogously, with cold tongues at the surface and low-potential-vorticity tongues aloft. *See* BAROCLINIC FIELD.

Other instabilities. Not all dynamic instabilities can be understood in the same terms in which barotropic and baroclinic stabilities may be comprehended. One such phenomenon is the Kelvin-Helmhotz instability, which occurs in stratified shear flows, that is, flows which have horizontal wind changing with height and which are at the same time convectively stable (θ increasing with height). If the wind shear is stronger than a critical value that depends on the degree of stratification, instability can occur. The resulting motions take the form of waves that often break, leading to turbulence. The Kelvin-Helmholtz instability is thought to be responsible for much of the clear-air turbulence encountered by aircraft. *See* CLEAR-AIR TURBULENCE.

There are still other forms of dynamic instability about which little is known. An example is fluid flow between two rigid, parallel plates moving at a constant speed with respect to each other. If the plates are moving slowly, the speed of the fluid flow varies smoothly from one plate to the other; this is known as Couette flow. Such a fluid has constant vorticity and is not stratified, so that it is not susceptible to any of the instabilities discussed previously. Yet when the plates are made to move fast enough, the fluid becomes unstable and quickly breaks down into a highly turbulent flow. Theoretical and laboratory work reported in 1986 has shown that flow within a rotating ellipse is unstable, even though it too has constant vorticity. Since perturbations to Couette flow have constant vorticity, the shape of the perturbation is usually elliptical and thus is unstable. This complex class of instabilities has begun to be explored, and the work may lead to important advances in the understanding of dynamic instabilities. *See* ATMOSPHERE; MICROMETEOROLOGY; NEWTONIAN FLUID.

Kerry A. Emanuel

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Dynamic meteorology

The study of those motions of the atmosphere that are associated with weather and climate. Atmospheric motions span an enormous range of spatial and temporal scales; dynamic meteorology concentrates mainly on large-scale and mesoscale motions. Large-scale motions are those with horizontal scales in excess of a few hundred kilometers and time scales longer than a day. Such motions are strongly influenced by the rotation of the Earth and by the vertical thermal stratification of the atmosphere. Mesoscale motions have horizontal scales in the range of a few kilometers to a few hundred kilometers; they are often associated with convective clouds and precipitation.

The mechanics and thermodynamics of the unsaturated atmosphere are governed by three fundamental conservation laws: (1) conservation of mass, expressed by the mass continuity equation; (2) conservation of momentum, expressed by Newton's second law of motion; and (3) conservation of thermodynamic energy, expressed by the first law of thermodynamics. For saturated conditions, conservation

of water substance must also be considered. *See* CONSERVATION OF MASS; CONSERVATION OF MOMENTUM; NEWTON'S LAWS OF MOTION; THERMODYNAMIC PRINCIPLES.

Statics. The vertical structure of the atmosphere is determined by the equation of state for an ideal gas and by the hydrostatic relationship. The former expresses the relationship among pressure, density, and temperature at any point; the latter expresses the balance between the upward-directed component of the pressure-gradient force (associated with the approximate exponential decrease of pressure with height) and the downward-directed gravity force. The equation of state and hydrostatic equation may be combined to form the hypsometric equation (1),

$$Z_2 - Z_1 = \frac{R}{g}\int_{p_2}^{p_1} T\, d\ln p \tag{1}$$

which relates the geopotential height differences ($Z_2 - Z_1$) between two pressure surfaces p_2 and p_1 to the mean temperature T in the layer between the two surfaces; and where R (the gas constant for dry air) $= 287\ \mathrm{J \cdot kg^{-1} \cdot K^{-1}}$ and g (the acceleration of gravity) $= 9.81\ \mathrm{m \cdot s^{-2}}$. Thus, pressure decreases more rapidly with height in cold air than in warm air. *See* GAS; HYDROSTATICS.

Except in regions of active precipitation, where the heating rate due to latent heat of condensation is large, temperature changes following the motion of individual parcels of air are controlled primarily by adiabatic expansion and compression as the air parcels move to lower or higher pressure. The thermodynamic state of such parcels can be characterized by the potential temperature θ, as in Eq. (2), where p_0 [$=10^5$ pascals (1000 millibars)] is

$$\theta = T\left(\frac{p_0}{p}\right)^{R/C_p} \tag{2}$$

a reference pressure and c_p ($= 1004\ \mathrm{J \cdot kg^{-1} \cdot K^{-1}}$) is the specific heat at constant pressure. The potential temperature is the temperature that a parcel of air at pressure p and temperature T would acquire if it were moved adiabatically to pressure p_0. When all diabatic heat sources can be neglected, θ remains constant in time for each air parcel. Normally, θ increases with altitude in the atmosphere, so that an air parcel displaced upward (downward) has a value of θ less (greater) than that of its environment, and hence experiences a net buoyancy force that tends to return it to its equilibrium level. The atmosphere is then said to be statically stable. When θ is constant with height, a condition that occurs when the temperature decreases with height at a rate $10^\circ\mathrm{C \cdot km^{-1}}$ ($30^\circ\mathrm{F \cdot mi^{-1}}$), the atmosphere is said to be neutrally stable; if θ decreases with height, the atmosphere is absolutely unstable and convective motion develops spontaneously.

If an air parcel is saturated, upward displacement causes water vapor to condense and release its latent heat of condensation; the potential temperature is then no longer conserved. It is possible, however, to define an equivalent potential temperature θ_e, as in Eq. (3), where L_c is the latent heat of condensa-

$$\theta_e = \theta \exp\frac{L_c q_s}{c_p T} \tag{3}$$

tion ($=2.5 \times 10^6\ \mathrm{J \cdot kg^{-1}}$ at 0°C or 32°F), and q_s is the saturation water-vapor mixing ratio (that is, the ratio of the mass of water vapor to mass of dry air for a saturated parcel). The equivalent potential temperature is conserved for both saturated and unsaturated motions provided that the temperature used in Eq. (3) is the temperature that the parcel would have if brought to saturation by adiabatic expansion. Because q_s decreases rapidly as temperature decreases, θ_e for saturated conditions generally decreases with height even when θ itself increases with height. The atmosphere is then said to be conditionally unstable since the static stability is positive for unsaturated parcel displacements but negative for saturated parcel displacements.

The hydrostatic relationship implies that pressure decreases monotonically with altitude. Pressure may thus be substituted for height as the independent vertical coordinate. If the atmosphere is everywhere statically stable so that θ increases monotonically with height, potential temperature can also be used as a vertical coordinate. Potential temperature coordinates are useful for analysis of adiabatic motions, since in that reference frame prediction of adiabatic flow is reduced to a two-dimensional problem of following the motion on θ surfaces. Isobaric coordinates, in which pressure is the vertical coordinate, have the advantage of eliminating any explicit reference to the density field. These are the most commonly used vertical coordinates in dynamic meteorology. *See* HYDROSTATICS; ISOBAR (METEOROLOGY).

Conservation laws in isobaric coordinates. The horizontal momentum equation, the mass continuity equation, and the first law of thermodynamics are expressed in isobaric coordinates as shown in Eqs. (4)–(6), where ∇ denotes the horizontal gradi-

$$\frac{d\mathbf{V}}{dt} = -f\mathbf{k} \times \mathbf{V} - g\nabla Z + \mathbf{Fr} \tag{4}$$

$$\nabla \cdot \mathbf{V} + \frac{\partial \omega}{\partial p} = 0 \tag{5}$$

$$\frac{dT}{dt} = S_p\omega + \frac{\dot{q}}{c_p} \tag{6}$$

ent evaluated at constant pressure, t denotes time, $\mathbf{V}$ is the horizontal velocity component, $f = 2\Omega \sin\phi$ is the Coriolis parameter (where $\Omega = 7.292 \times 10^{-5}\ \mathrm{s^{-1}}$ is the angular velocity of rotation of the Earth and ϕ is latitude), $\mathbf{k}$ is the vertical unit vector, Z is the height of a surface of constant pressure, $\mathbf{Fr}$ is the force of friction, $\omega = dp/dt$ is the vertical motion in isobaric coordinates, $S_p = -T\,\partial \ln\theta/\partial p$ is the static stability, and $\dot{q}$ is the heating rate.

Equations (4) and (6) are expressed in lagrangian form; that is, the notation d/dt expresses the rate of change with respect to time following the motion of an individual parcel. For analysis and prediction purposes, it is usually more convenient to use the

eulerian form in which time derivatives are computed at fixed points in space (as partial derivatives). The differential operators for the lagrangian and eulerian forms are related by Eq. (7), where the second

$$\frac{d}{dt} = \frac{\partial}{\partial t} + \mathbf{V} \cdot \nabla \tag{7}$$

term on the right side expresses the change due to advection by the horizontal wind.

Equation (4) states that the rate of change of the horizontal velocity following the motion of an air parcel is due to the net imbalance among three forces: the Coriolis force due to the rotation of the Earth, the horizontal pressure-gradient force, and the force due to frictional dissipation (mostly at the surface of the Earth). Equation (5) states that the divergence of the horizontal velocity, which tends to stretch or shrink the horizontal cross section of an air parcel, must be balanced by vertical shrinking or stretching due to the change in vertical mass flow with altitude. Equation (6) states that the rate of change of temperature following the motion is due to the sum of the adiabatic expansion or compression due to vertical motion, plus the net heating due to such heat sources as condensation and solar and thermal radiation (generally referred to as diabatic heating). Equation (2) is an integral of Eq. (6) in the case where diabatic heating vanishes.

Balanced flow. When the diabatic heating rate $\dot{q}$ and the frictional force **Fr** can be neglected, Eqs. (4)–(6), together with the differential form of the hypsometric equation (1), form a closed set for prediction of $\mathbf{V}$, ω, Z, and T subject to suitable initial and boundary conditions. This is a complicated set of nonlinear partial differential equations that in general must be solved by numerical methods. However, for large-scale atmospheric motions certain approximate solutions can provide considerable insight.

If friction can be ignored, which is often approximately the case above the lowest kilometer of the atmosphere, large-scale extratropical motions are approximately in geostrophic balance; that is, the horizontal pressure-gradient force and the Coriolis force are nearly equal and opposite. The wind velocity for which this balance is exact is referred to as the geostrophic wind, and is determined by Eq. (8), a

$$f\mathbf{V}_g = g\mathbf{k} \times \nabla Z \tag{8}$$

special case of Eq. (5). The force balance for the geostrophic wind is illustrated in **Fig. 1**. The Coriolis force acts perpendicular to the wind direction (to the right of the wind in the Northern Hemisphere), while the pressure-gradient force is directed down the gradient of geopotential height. Thus, geostrophic balance requires that the wind be parallel to the height contours with speed proportional to the height gradient at constant pressure. *See* CORIOLIS ACCELERATION; GEOSTROPHIC WIND.

Equations (8) and (1) may be combined to yield the thermal wind equation (9), where $\mathbf{V}_T =$

$$f\mathbf{V}_T = g\mathbf{k} \times \nabla(Z_2 - Z_1) = R\int_{p2}^{p_1} (\mathbf{k} \times \nabla T) d \ln p \tag{9}$$

$\mathbf{V}g(p_2) - \mathbf{V}g(p_1)$. Equation (9) states that the vector difference in the geostrophic wind velocity between two pressure surfaces is proportional to the horizontal gradient of the mean temperature in the layer between the two surfaces. This relationship, which is a consequence of geostrophic and hydrostatic balance, provides a strong constraint on the horizontal velocity and temperature distributions. For motions that are nearly geostrophic and nearly hydrostatic, the wind and temperature fields cannot vary independently, but they are closely coupled through their mutual relationship to the temperature (or density) distribution. *See* THERMAL WIND.

Vorticity and potential vorticity. Equation (4) indicates that momentum is not conserved on a fluid parcel, even when the flow is adiabatic and frictionless; its rate of change depends on the small difference between the Coriolis force and the pressure-gradient force. A dynamical quantity that is conserved following the motion under such conditions is the potential vorticity. Potential vorticity can be expressed in its simplest form by using potential temperature coordinates, as is shown in Eq. (10), where $\zeta = \mathbf{k} \cdot (\nabla \times \mathbf{V})$

$$P = -(\zeta + f)\frac{\partial \theta}{\partial p} \tag{10}$$

is the vertical component of the relative vorticity and f, the Coriolis parameter, is the vertical component of the Earth's vorticity.

Equation (10) indicates that the product of the absolute vorticity ($\zeta + f$), which is a measure of the rotation of a fluid particle about a local vertical axis in inertial space, and the stability parameter $-\partial\theta/\partial p$, which is a measure of the vertical extension of the fluid parcel, is conserved following the motion. Potential vorticity conservation is the fluid dynamic analog of angular momentum conservation in solid body mechanics. As an air parcel moves from a region where it has weak static stability (that is, where

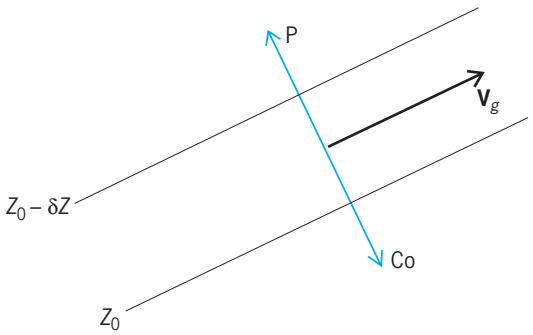

Fig. 1. Force balance for the geostrophic wind. P = pressure-gradient force. Co = Coriolis force. Z_0 and $Z_0 - \delta Z$ are constant-height contours on an isobaric surface. $\mathbf{V}_g$ = geostrophic wind.

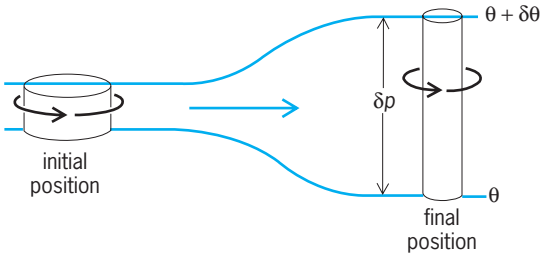

Fig. 2. Cylindrical column of air moving adiabatically, conserving potential vorticity. θ = potential temperature. p = pressure. (*After J. R. Holton, An Introduction to Dynamic Meteorology, 3d ed., Academic, 1992*)

$-\partial\theta/\partial p$ is small) to a region where its static stability is strong, the parcel must shrink in vertical extent and increase in horizontal area, with a resulting decrease in absolute vorticity (**Fig. 2**). It turns out that with specifications of suitable boundary conditions, the distribution of the large-scale wind and temperature fields in three-dimensional space can be uniquely determined from the distribution of potential vorticity. Since potential vorticity tends to be conserved following the motion, the evolution of the distribution of potential vorticity in time can be easily predicted and utilized to infer the evolution of wind and temperature. *See* DYNAMIC INSTABILITY.

Planetary boundary layer. Within the lowest kilometer of the atmosphere, there is a boundary layer in which the frictional force due to small-scale turbulent motions represents an important contribution to the force balance. The structure of the atmospheric boundary layer depends strongly on the vertical stratification in the layer. In regions of strong static stability (where potential temperature increases rapidly with altitude), turbulence is suppressed and the frictional force is weak except in the lowest several meters above ground, where strong vertical shear of the wind provides a mechanical source for turbulence. When there is strong surface heating (as over deserts) the potential temperature may actually decrease with height near the ground. Boundary-layer turbulence is then convectively driven, and the frictional force may be strong throughout at least the lowest kilometer of the atmosphere. Over vast areas of the Earth's surface, the boundary layer is typically near neutral stability. In that case, it is permissible to approximate the frictional force as a drag proportional to the wind velocity: $\mathbf{Fr} = -k\mathbf{V}$, where k^{-1} is the time scale for friction to deplete the velocity by a factor of e^{-1}. Typically, $k \simeq 10^{-5}$ per second, and the force balance in the boundary layer is a three-way balance among the friction force, Coriolis force, and pressure-gradient force.

This balance is shown schematically in **Fig. 3**. Since the Coriolis force always acts to the right of the wind direction in the Northern Hemisphere, and the frictional force is opposite to the wind direction, a balance can be achieved only if the wind has a component directed across the pressure field toward lower pressure. This component produces a net frictional convergence into surface low-pressure systems, and divergence out of surface highs. *See* CONVECTIVE INSTABILITY.

Atmospheric waves. The dynamics equations (4)–(6) are capable of describing a wide variety of systems of atmospheric motion. In order to classify and analyze these, it is useful to divide the horizontal velocity into rotational and irrotational parts. The circulation in large-scale weather systems is dominated by the rotational part of the wind; outside the tropics this can be approximated by the geostrophic wind. The irrotational part of the velocity field dominates in small-scale atmospheric motions in which the divergence of the horizontal wind plays a crucial role. Although many of the most important features are inherently nonlinear, there are several types of wave solutions that can be analyzed by using linearized versions of the dynamics.

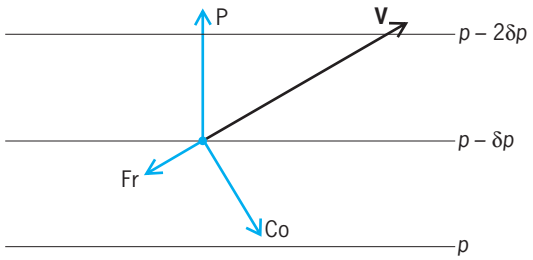

Fig. 3. Balance of forces in a well-mixed planetary boundary layer. P = pressure-gradient force. Co = Coriolis force. Fr = frictional force. p = pressure. (*After J. R. Holton, An Introduction to Dynamic Meteorology, 3d ed., Academic, 1992*)

Buoyancy and inertial waves. Waves in fluids are motions in which information can be transferred (usually by pressure perturbations) without material transport of air parcels. In all cases, atmospheric waves result from a balance between the fluid inertia and some restoring force. For example, the vertical stratification of the atmosphere causes a parcel displaced vertically to experience a restoring force due to its buoyancy; the resulting oscillation, which is dominated by the divergent part of the wind field, is referred to as a buoyancy (or internal gravity) wave. In a somewhat analogous fashion, the lateral deflection due to the Coriolis force causes a parcel displaced horizontally to experience a lateral restoring force; the resulting oscillation is known as an inertial wave. Waves in which both buoyancy and Coriolis forces are important are known as inertia-gravity waves.

The longitude–height structure of an eastward propagating buoyancy wave is shown in **Fig. 4**. Phase lines tilt toward the east with height, so that as the wave passes a particular location there is an apparent downward propagation in time. However, the energy flux in this case is actually directed upward, parallel to lines of constant phase. Buoyancy waves provide an important mechanism for transferring momentum from the lower atmosphere into the upper atmosphere. The buoyancy wave frequency ν is given by Eq. (11), where $N = (g\partial \ln \theta/\partial Z)^{1/2}$ and

$$\nu = N\cos\alpha \qquad (11)$$

α is the angle of the phase lines to the local vertical. The maximum possible frequency N thus occurs for vertically oriented phase lines. In the lower atmosphere, N typically corresponds to a period of about 10 min. Only those waves in which the ratio of horizontal to vertical scale is very large have frequencies sufficiently low to be of much meteorological importance.

An important class of low-frequency buoyancy waves comprises lee waves, formed when air is forced to flow over a mountain barrier. The oscillations that are excited by such forcing may continue many wavelengths downstream of the mountain barrier. When sufficient moisture is present, condensation may occur in the updraft portions of the waves and regularly spaced bands of cloud may form. In this case, the waves are stationary with respect to the ground, but propagate relative to an observer moving with the mean wind. *See* MOUNTAIN METEOROLOGY.

Planetary waves. The most important class of large-scale atmospheric waves comprises planetary waves. They are characterized by oscillations in the rotational part of the horizontal wind. In these waves the latitudinal gradient of the mean atmospheric potential vorticity provides the restoring force that acts to oppose meridional parcel displacements. The simplest example of a planetary wave occurs in a barotropic atmosphere (an atmosphere in which potential temperature is a function of pressure alone). In a barotropic atmosphere, conservation of potential vorticity [see Eq. (10)] reduces to conservation of absolute vorticity following the motion described in Eq. (12).

$$\frac{d(\zeta + f)}{dt} = 0 \tag{12}$$

If a chain of air parcels extending around a latitude circle with $\zeta = 0$ at time t_0 is given a small meridional displacement δy, then from application of the conservation law in Eq. (12), at time t_1 the relative vorticity will be as shown in Eq. (13), where $\beta = df/dy$,

$$\zeta_{t_1} = f_{t_1} - f_{t_0} = -\beta\ \delta y \tag{13}$$

the rate of increase of the Coriolis parameter with latitude. From Eq. (13) it is evident that the relative vorticity perturbation will be positive for a southward displacement and negative for a northward displacement. Since positive (negative) relative vorticity corresponds to counterclockwise (clockwise) rotation, the induced perturbation velocities will tend to move the displacement pattern toward the west, as illustrated in **Fig. 5**. Thus, the longitudinal propagation of planetary waves is always westward relative to the mean wind. Such waves are transverse in the sense that the fluid parcel oscillations, which are in the north-south direction, are perpendicular to the direction of wave propagation.

Quantitative analysis reveals that planetary waves are highly dispersive; that is, their longitudinal phase speeds are strongly dependent on wavelength. In cartesian coordinates this dependence, in the presence of a mean zonal flow $\overline{u}$, has the form shown in Eq. (14), where $c_x - \overline{u}$ is the phase speed relative

$$c_x - \overline{u} = -\frac{\beta}{4\pi^2}\left[\frac{1}{L_x^2} + \frac{1}{L_y^2}\right]^{-1} \tag{14}$$

to the mean flow, and L_x and L_y are the wavelengths in longitude and latitude, respectively. Equation (14) shows that planetary waves always propagate westward relative to the mean flow, with phase speed proportional to the square of the wavelength. For waves corresponding to typical extratropical storms with wavelengths of a few thousand kilometers, the planetary wave phase speed is typically less than 10 m · s^{-1} (33 ft · s^{-1}), which is less than the mean wind speed. Thus, relative to the ground, such

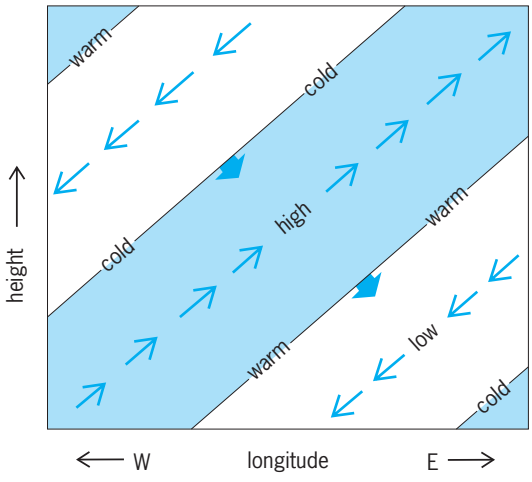

Fig. 4. Longitude–height structure of an eastward-propagating buoyancy wave. Thin arrows indicate the perturbation velocity field; heavy arrows indicate the phase velocity. The labels "high" and "low" and "warm" and "cold" indicate the axes of maximum perturbation pressure and temperature, respectively.

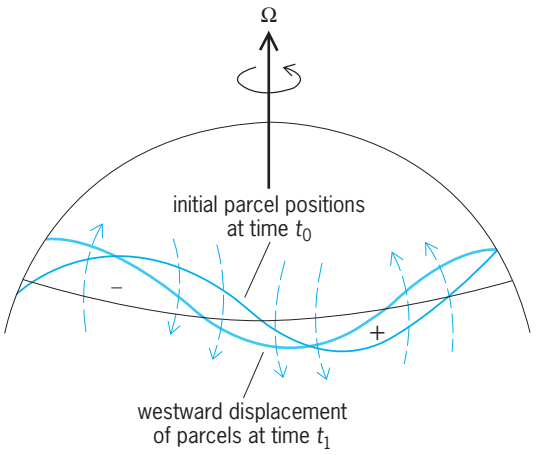

Fig. 5. Propagation mechanism for a planetary wave. Perturbation vorticity is shown by + and −, induced velocity field is shown by broken arrows. Ω = angular velocity of rotation of the Earth. (*After J. R. Holton, An Introduction to Dynamic Meteorology, 3d ed., Academic, 1992*)

disturbances tend to move eastward, but at a speed less than the average eastward wind speed.

Equation (14) also indicates that planetary waves that are stationary relative to the ground ($c_x = 0$) can exist only when the mean zonal flow is positive (that is, from the west). Thus, stationary waves are excited in midlatitudes when westerly winds encounter mountain barriers. It is this process that accounts for the lee side trough observed to the east of the Rocky Mountains.

Baroclinic instability. The barotropic planetary waves described above cannot represent any motions that involve conversion of energy between potential and kinetic forms. Such energy conversion can occur only in the presence of baroclinity, that is, variations of potential temperature on isobaric surfaces. Baroclinic energy conversion processes are responsible for the growth and maintenance of most large-scale weather disturbances.

In midlatitudes in the troposphere, potential temperature normally decreases from the Equator to the pole on isobaric surfaces. This decrease does not occur uniformly, but tends to be concentrated in the jet stream, a narrow band of strong westerly winds in the upper troposphere that encircles the globe in midlatitudes. The association of the jet stream with a strong meridional temperature gradient is a consequence of the thermal wind relationship, Eq. (9), which states that westerly winds must increase rapidly with height where the temperature decreases rapidly in latitude. *See* JET STREAM.

When the shear of the zonal wind is sufficiently strong that the meridional gradient of potential vorticity on a constant potential temperature surface is locally reversed, or when there is a nonvanishing gradient of potential temperature at the surface of the Earth, the linearized equations have solutions in the form of exponentially growing disturbances. These baroclinically unstable modes have growth rates, structures, and scales similar to those observed in developing extratropical cyclones. They are quasigeostrophic in the sense that the geostrophically balanced rotational component of the velocity field strongly dominates over the divergent component. The latter is, however, crucial in the energy cycle through which the disturbances convert potential energy stored in the meridional gradient of potential temperature into disturbance kinetic energy. Baroclinic instability provides a mode whereby infinitesimal disturbances may amplify into large-amplitude storms. However, in many cases it appears that storms in the atmosphere grow through nonlinear interactions involving preexisting disturbances. Such processes must generally be studied by numerical simulations on high-speed computers.

Mesoscale convective systems. For horizontal scales less than several hundred kilometers, the major energy source is not baroclinic instability; it is latent heat release by cumulonimbus clouds. The convective storms associated with such clouds can occur only when the atmosphere is conditionally unstable [Eq. (2)], sufficient moisture is present, and there is an initial disturbance strong enough to lift air parcels high enough to release the conditional instability. Mesoscale convective systems take a variety of forms. Among these are hurricanes, squall lines, and mesoscale convective complexes. In all of these systems the release of latent heat by convective clouds is the primary energy source, but the character of the large-scale environmental flow is generally important for determining the mode of organization for the convection. There is no single theory, analogous to the theory of baroclinic instability, that can explain the origin and growth of mesoscale convective systems. Most studies of such systems have involved numerical modeling of particular cases. *See* HURRICANE; MESOMETEOROLOGY; SQUALL LINE.

Numerical weather prediction. The dependent variables in the dynamics equations (4)–(6) are a set of fields that are continuous functions of space and time. In order to solve these equations numerically, it is necessary to approximate the various fields, such as velocity, temperature, and pressure, either by their values on a finite set of grid points in three-dimensional space (the finite difference method) or by truncated series of orthogonal functions (the spectral method). In current global weather prediction models, a mixture of techniques is often employed. Nearly all models use finite difference representations for the vertical coordinate and time; the horizontal variation is generally represented either by a network of grid points at uniform intervals of latitude and longitude or by a finite set of spherical harmonics. In the latter representation, a field, such as temperature, at a given pressure and time is given by a summation of the form shown in Eq. (15),

$$T(\lambda, \phi, p, t) = \sum_{m=-J}^{+J} \sum_{n=|m|}^{|m|+J} T_n^m(p, t) P_n^m(\sin\phi) \exp(im\lambda) \qquad (15)$$

where λ and ϕ denote longitude and latitude, respectively, and $P_n^m(\sin\phi)$ designates the associated Legendre function of the first kind. For a given spectral mode, the integer index m represents the number of waves around a latitude circle, while $n - m$ is the number of waves between the north and south poles, and J is the total number of modes retained in the expansion. The spectral method generally provides higher accuracy, especially in predicting the speed of motion of short wave features, than can be obtained with a finite difference representation of similar resolution. *See* LEGENDRE FUNCTIONS; SPHERICAL HARMONICS.

Because the lower boundary of the atmosphere does not correspond to an isobaric surface, pressure is not a very convenient vertical coordinate for numerical prediction models. In practice, therefore, pressure is usually replaced by the coordinate $\sigma = p/p_s$, in which pressure is normalized by the local surface pressure p_s. Typically, about 10–15 levels are specified in the interval $\sigma = 1$ to $\sigma = 0$, and the horizontal resolution is of order 3–5° in longitude and latitude.

To predict the weather dynamically, it is necessary to solve the dynamics equations by integrating in time, starting from an initial state of the atmospheric variables determined from observations. In practice, observational errors and poor data coverage, particularly over the oceans, make it impossible to exactly determine the initial state of the atmosphere. Moreover, the state determined from observations generally does not properly represent the true dynamical balance among the pressure and wind fields. This imbalance introduces noise that is interpreted as high-frequency inertia-gravity waves. If a prediction is attempted from such an initial state, the noise rapidly dominates the true, slowly evolving weather disturbances. In order to prevent the growth of such spurious noise, the analysis must be initialized by processing it in a manner that assures a dynamical balance between the pressure and wind fields while preserving the actual observations as closely as possible.

However, even if the initial state were known perfectly, there still would be a limit beyond which errors would dominate the forecast. This predictability limit arises because the atmosphere has a continuous spectrum of scales of motion. There are inevitably scales too small to be resolved by the model, and through nonlinear interactions these would eventually come to dominate the forecast. Thus, the theoretical length of useful forecasts for transient weather disturbances is about 3 weeks. The predictive skill of present models is still far short of this limit. *See* WEATHER FORECASTING AND PREDICTION.

Global climate modeling. In addition to their role in weather prediction, dynamical models can also be used to simulate global climate. Climate is the study of the average state of the atmosphere and its seasonal and interannual variability. Climate is determined by the joint influence of energy sources and sinks at the Earth's surface, and the transformation and transport of energy in the atmosphere and the oceans. Models that simulate these processes are usually referred to as general circulation models. Such models must contain accurate representations of all the important physical processes that influence the circulation.

Among these processes are heating and cooling by solar and infrared radiation; vertical transfer of heat and momentum by convective clouds; latent heating due to evaporation, condensation, and freezing of water; and drag and mixing due to small-scale turbulent motions in the boundary layer. In most cases, these processes involve scales of motion smaller than those explicitly resolved by the global model. They must then be parametrized in terms of the resolved variables. It is often not clear how to accomplish this objective. Thus, parametrization of subgrid-scale fields is an area of active research for global modeling. *See* ATMOSPHERE; ATMOSPHERIC GENERAL CIRCULATION; CLIMATE MODELING; FLUID-FLOW PRINCIPLES; METEOROLOGY; WIND.

James Holton

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Dynamic nuclear polarization

The creation of assemblies of nuclei whose spin axes are not oriented at random, and which are in a steady state that is not a state of thermal equilibrium. Under commonly occurring conditions, the spin axes of nuclei (with nonzero spin) are oriented at random; where this is not so, the nuclei are said to be polarized. Assemblies of polarized nuclei are not in a state of thermal equilibrium except under rather extreme conditions (for example, temperatures below 10 millikelvins or 0.02°F above absolute zero, and magnetic fields greater than several teslas), and schemes have been devised to produce polarized assemblies, in a steady state which is not a state of thermal equilibrium, under less extreme conditions. Such schemes constitute dynamic nuclear polarization.

Among the many applications of polarized nuclei are the following. Nuclear forces are spin-dependent, and although the spin-dependent part can be found by using unpolarized assemblies, the experiments are simpler and their interpretation is clearer if polarized nuclei are used. Assemblies of polarized nuclei have a lower geometrical symmetry than assemblies of randomly oriented nuclei, and so these have been used to investigate the fundamental symmetries of nature. Polarized nuclei have been used to enhance the signal in free precession magnetometers and similar instruments, and the use of an assembly of polarized nuclei as a gyroscope has also been suggested. *See* MAGNETOMETER; NUCLEAR ORIENTATION; PARITY (QUANTUM MECHANICS); SPIN (QUANTUM MECHANICS).

The simplest and most common parameter used to define the state of polarization is called the "polarization" or vector polarization $\mathbf{P}$. It is a vector whose component along any given axis (for example, the $z$axis) is the average value of the projection on that axis of a unit vector parallel to the nuclear spin axis. Often, nuclei lie with their spins pointing along the $z$axis in equal numbers in both directions, but none lies in the x-y plane. Such a system is not random, but $\mathbf{P} = 0$. The parameter used to quantify this arrangement is the average value of $(^3/_2 \cos^2 \Theta - {}^1/_2)$ and is one component of the so-called tensor polarization. In general, the average value of any zonal harmonic is a measure of a kind of polarization. It is a property of nuclear spin operators that harmonics of order greater than $2J$ vanish identically, where J is the nuclear spin. In particular, spin 0 nuclei cannot be polarized and spin $^1/_2$ nuclei can have a vector polarization only. *See* ANGULAR MOMENTUM; SPHERICAL HARMONICS.

There are, in general, two common features of all dynamic nuclear polarization schemes. There is

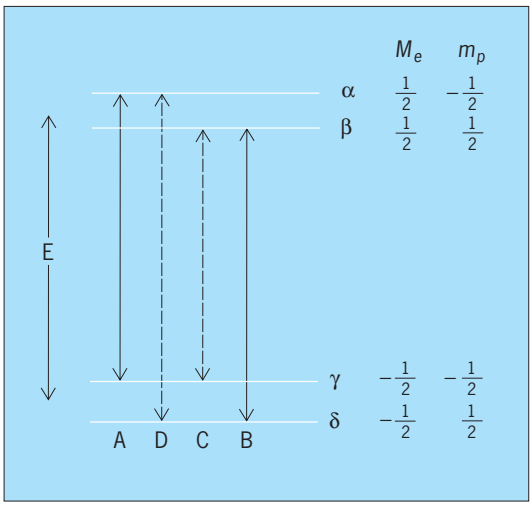

Fig. 1. Energy levels of an electron and a proton in a magnetic field.

some external system or interaction, for example, an oscillating magnetic field, which disturbs the system; this is called the pumping mechanism. There are also processes which tend to restore the system to thermal equilibrium; these are called relaxation processes. However, the functions of these two processes are not easily distinguished, for, in many schemes, the relaxation mechanisms actually play a part in setting up the nuclear polarization, this being one of the stages back to thermal equilibrium.

Microwave pumping. Consider an electron spin and a nucleus of spin $^1/_2$ (for example, a proton) in close association. Typically, the electron spin is part of a paramagnetic ion, and the proton is a hydrogen nucleus in a nearby water molecule. In a magnetic field of about 1 T, the energy levels of this combination are as shown in **Fig. 1**. M_e and m_p are, respectively, the electron and proton spin magnetic quantum numbers which specify whether the electron and proton spins are oriented in the same direction as the magnetic field or in the opposite direction. An oscillating magnetic field is applied to induce transitions between pairs of levels. The two transitions A and B, in which only an electron spin is reversed, constitute the "allowed" transition. The two weaker transitions, C and D, in which a nuclear spin is reversed, occurring one at a higher and the other at a lower field (or frequency) than the allowed transition, are called forbidden transitions. *See* MAGNETIC RESONANCE; SELECTION RULES (PHYSICS).

In the most common and most successful mechanism of this type for polarizing protons, one of the forbidden transitions (C, for instance) is excited strongly enough to saturate it. Then the populations of levels β and γ are equal, and are kept equal by the oscillating magnetic field. The strongest relaxation processes, which try to bring the system into thermal equilibrium, are associated with the strongest transitions A and B. As a result of these processes, fewer systems occupy level α than level γ, and similarly, fewer systems occupy level β than level δ. Thus more proton spins are oriented in one direction ($m_p = +^1/_2$) than in the opposite direction ($m_p = -^1/_2$), and a nuclear polarization results. By saturating the other forbidden transition, an equal and opposite polarization results. *See* MAGNETIC RELAXATION.

In another method the allowed transition is saturated, and so the populations of levels α and γ are equal, and similarly the populations of levels β and δ are equal. The strong relaxation processes, which involve only the inversion of an electron spin, are ineffective since they are swamped by the radio-frequency field. The "cross relaxation" processes, which connect level α with level δ (called the flip-flop relaxation), and which connect level β with level γ (called the flip-flip relaxation), are of different strengths. This is because in the flip-flop transition, spin angular momentum is conserved, but not so in the flip-flip transition. The relative populations of the levels are thus determined by the relative strengths of the two processes. If, as in the case of metals, the flip-flip transition can be neglected, the flip-flop process results in fewer systems occupying level α than level δ, and a nuclear polarization results. If the electron spin and the nuclear spin are coupled by magnetic dipole interaction, and direct nuclear spin lattice relaxation can be neglected, the flip-flip process is twice as strong as the flip-flop process, and a smaller nuclear polarization of the opposite sign results.

Although this explanation describes very effectively what happens, it is oversimplified. Nuclei do not interact in turn with individual electron spins, and the latter are not independent. Especially when protons are polarized in organic materials, the effect of the oscillating magnetic field is to cool the assembly of interacting electron spins, which then cool the nuclei by thermal contact. These processes are known by the acronym DONKEY effects. *See* PARAMAGNETISM.

In order to produce large polarizations, the external magnetic field should be of the order of 2 T, the frequency of the oscillating field should be about 70 GHz, and the temperature should be 1 K (2°F above absolute zero) or less. Values of polarization of 0.8 or more (corresponding to 90% of the spins oriented in one direction and only 10% in the opposite direction) are regularly obtainable for protons polarized in $La_2Mg_3(NO_3)_{12} \cdot 24H_2O$ in which 1% of the lanthanum, La, is replaced by neodymium, Nd; in an empirically determined mixture consisting mainly of butanol, C_4H_9OH, and about 1% of porphyrexide, a free radical; or in ethylene glycol or propanendiol with about 1% of an organic Cr^V complex. Deuteron polarizations of 0.4 have been obtained by using C_4D_9OD and porphyrexide at temperatures of 0.3 K (0.5°F above absolute zero). *See* COORDINATION CHEMISTRY; COORDINATION COMPLEXES.

Butanol and diol systems are favored to make targets of polarized protons for experiments in high-energy physics. The beam passing through the target causes radiation damage which adversely affects thepolarization. To a first approximation, the

polarization P falls off as an exponential function of the dose ϕ, given by the equation below, where

$$P = P_0 \exp(-\phi/\phi_0)$$

P_0 is the initial polarization and ϕ_0 is approximately 2.5×10^{15} minimum ionizing particles per square inch (4×10^{14} minimum ionizing particles per square centimeter); but this levels off to a constant value of approximately $0.7P_0$. This effect is caused by the production of free radicals, and free electrons which are trapped to form color centers, both of which are paramagnetic. Most of the radiation damage can be annealed out by raising the temperature of the target material to about −208°F (140 K) for butanol, or −136°F (180 K) for diols, and keeping it there for about 15 min. *See* COLOR CENTERS; FREE RADICAL.

Polarization by rotation. This method was most spectacularly demonstrated by rotating a crystal of yttrium ethylsulfate, $Y(C_2H_5SO_4)_3 \cdot 9H_2O$, in which a few percent of yttrium, Y, was replaced by ytterbium, Yb, in a magnetic field of 1 T at a temperature of 1.4 K (2.5°F above absolute zero), at a rate of 60 revolutions per second. The proton polarization was found to be 0.19 (corresponding to 59.5% of the spins oriented in one direction and 40.5% in the other direction); higher values can be obtained by modifying the experimental conditions. The splitting of the two energy levels of the protons in the external magnetic field is independent of the orientation of the crystal, but the splitting of the two energy levels of the Yb^{3+} ion depends strongly on the orientation of the magnetic field relative to the c axis of the crystal, being quite large when these are parallel and almost zero when they are perpendicular. The Yb^{3+} ions are strongly coupled to the lattice when this angle is 45°, but almost isolated when it is 0 or 90°. Thus, as the crystal is rotated, when it reaches the 45° position, the Yb^{3+} ions quickly come to equilibrium at the temperature of the lattice. Then, as the crystal is rotated to the 90° position, the energy levels of the Yb^{3+} ions move together; this is equivalent to adiabatic demagnetization, and the spin temperature of the Yb^{3+} ions drops. Near the 90° position, the Yb^{3+}ions can exchange energy with the protons, because the two splittings are the same, and a flip-flop transition involving a Yb^{3+} ion and a proton conserves energy. The protons and the Yb^{3+} ions thus come to thermal equilibrium at the temperature of the Yb^{3+}ions after demagnetization. Continuous rotation repeats the cycle, and the ideal steady state is one in which the proton polarization is the same as the polarization of the Yb^{3+} spins at the 45° position. *See* ADIABATIC DEMAGNETIZATION; MAGNETIC SUSCEPTIBILITY.

Optical pumping. This method, pioneered by Alfred Kastler and Jean Brossel, makes use of the fact that circularly polarized light carries angular momentum, and when it is absorbed, that angular momentum is given to the absorber. Typically circularly polarized resonance radiation (such as sodium-D radiation) is incident on, for example, sodium vapor at room temperature (at a pressure of about 10^{-6} torr or 10^{-8} lb/in.^{2} or 10^{-4} pascal) in a small magnetic "guide" field directed along the direction of the light beam. It excites only the σ^+ transitions in which the magnetic quantum number of the absorbing atom, which specifies its angular momentum along the direction of the magnetic field, increases by 1. When the excited state decays, on the average there is no change in the magnetic quantum number. Thus an equilibrium is set up in which the ground-state atom spins are polarized. Polarizations of almost 1.0 can easily be obtained for the alkali metals. *See* POLARIZED LIGHT.

The absorbing atoms are usually mixed with a gas such as argon, called a buffer gas, at a pressure of 1–100 torr (0.01–1 lb/in.^{2} or 10^2–10^4 Pa) to delay their diffusion to the walls of the container where they may be depolarized. If some other gas is present instead of, or in addition to, the buffer gas, the angular momentum of the absorber is transferred in collisions to the atoms of this gas and can end up as a nuclear polarization.

This method has been very successfully applied to ^{3}He, which acts as a buffer gas. The absorbing atoms are metastable 2^3S_1 ^{3}He atoms produced by striking a weak electrodeless discharge in the ^{3}He gas. Polarizations of up to 0.4 have been achieved; under suitable conditions, when the polarizing mechanism is turned off, the polarization decays with a time constant of the order of several days. *See* OPTICAL PUMPING.

James M. Daniels

Laser-induced polarization. Nuclear orientation through optical pumping can also be achieved by using lasers. This method has been refined with the availability of high-power tunable continuous-wave dye lasers. The notion of a preferred direction in space is passed on to an ensemble of randomly oriented nuclei through atomic transitions induced by a polarized laser beam.

The energy levels in an atom exhibit fine and hyperfine structure. The hyperfine structure arises through the interaction of the magnetic and quadrupole moment of the nucleus with the atomic electrons. The hyperfine energy levels are classified in terms of the total angular momentum of the electron-nuclear system $\vec{F} \cdot \vec{F} = \vec{I} + \vec{J}$, where $\vec{I}$ and $\vec{J}$ are the electron and nuclear angular momenta respectively. $\vec{F}$ can take the smaller of $2I + 1$ or $2J + 1$ values. The projection of $\vec{F}$ on an arbitrary (quantization) axis, labeled M_F, can take on the values in the range $-F$ to F in integral steps. In the absence of electromagnetic fields, the various M_F levels associated with a given$F$ have the same energy and, under normal conditions in a large ensemble of atoms, are equally likely to be populated. Optical transitions can be induced by using a polarized laser to alter the random distribution of population among the M_F states; this constitutes nuclear orientation. The transitions are governed by selection rules. For electric dipole transitions between states of opposite parity, F can change either by $+1$, -1, or 0 (transitions between a state $F = 0$ and another $F' = 0$ are not allowed). M_F can change by $+1$, -1, or 0. The M_F selection rule

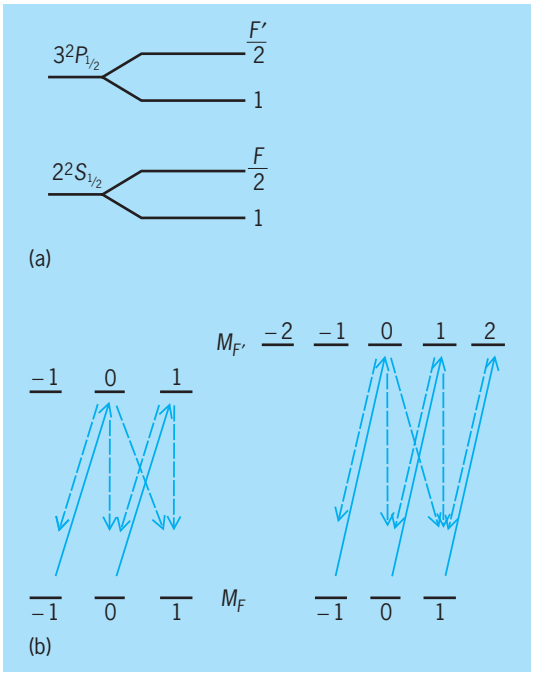

Fig. 2. Optical pumping of sodium-23. (*a*) Fine and hyperfine levels in the ground and first excited state of ^{23}Na. (*b*) An optical pumping cycle showing the allowed transitions, for excitation with a σ^+ polarized laser beam, from and to the $F = 1$ hyperfine level in the ground state. For the excitation, M_F can change by only +1, and for the decay, M_F can change by +1, −1, or 0.

depends further on the polarization of the laser light. *See* HYPERFINE STRUCTURE; NUCLEAR MOMENTS.

As an example, consider the optical pumping of a sodium-23 (^{23}Na) atom using σ^+ (circularly) polarized laser light. The quantization axis is chosen as the direction of propagation of the laser beam. For the ground state of ^{23}Na, $I = {}^3/_2$, $J = {}^1/_2$, and $F = 1, 2$. For the first excited state of opposite parity, $I = {}^3/_2, J = {}^1/_2$, and $F' = 1, 2$. These levels are shown in **Fig. 2***a*. The allowed transitions ($\Delta M_F = 1$) from the $F = 1$ ground state to the $F' = 1, 2$ excited states are shown in Fig. 2*b* as solid arrows. Transitions from and to the $F = 2$ level have been omitted for clarity. The spontaneous decay paths from the excited states have been shown as broken arrows. This process of excitation and decay constitutes an optical pumping cycle. A cursory inspection of Fig. 2*b* reveals that transitions populating the $M_F = 1$ level far outnumber those depleting it. After relatively few such pumping cycles the population in the atomic ensemble is found predominantly in the higher M_F levels. Under these conditions the spins of most of the nuclei are oriented in the direction of the laser beam. Such a technique has been used successfully to produce several species of nuclei with essentially all of the nuclear spins oriented in the same direction. Nuclear orientation can also be achieved by using a linearly polarized laser beam.

The extent to which nuclei can be oriented in this manner is limited by several factors. Some of the factors limiting the efficiency of an optical pumping cycle are: how well the laser bandwidth is matched to the absorption profile of the transitions being pumped; the oscillator strengths of the atomic transitions; the length of the laser-atom interaction time; and excited and ground-state relaxation through atom-atom and atom-wall collisions and through precession in stray magnetic fields. The thrust of research in the field of laser-induced nuclear orientation has been in overcoming these difficulties. *See* ATOMIC STRUCTURE AND SPECTRA; LASER; ZEEMAN EFFECT.

B. Shivakumar

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Dynamic similarity

A relationship existing between two fluid flows when they have identical types of forces that are parallel at all corresponding points, with magnitudes related by a constant scale factor. Dynamic similarity makes it possible to scale results from model tests to predict corresponding results for the full-scale prototype.

Dynamic similarity requires faithful reproduction of detail on the model (geometric similarity); the same flow pattern, including boundary shapes (kinematic similarity); and test conditions that match relevant dimensionless ratios between model and prototype. Dynamically similar flows are said to be homologous.

A rigorous method to find the dimensionless ratios is to nondimensionalize the governing equations and boundary conditions. The dimensionless coefficients are duplicated to obtain dynamic similarity. A practical method is to list the parameters and form dimensionless ratios by dimensional analysis.

Dimensionless parameters simplify presentation of experimental results. An example is the pressure coefficient given by the equation below, where p,

$$C_p = \frac{(p - p_{\text{ref}})}{{}^1/_2 \rho V^2}$$

p_{ref}, ρ, and V are local pressure, reference pressure, density, and velocity, respectively. Corresponding points in dynamically similar flows have equal pressure coefficients even when fluid density or speed differs. Hundreds of dimensionless parameters have been derived and named. A few parameters are in especially frequent use (see **table**), and they may be combined to obtain other parameters.

It may not be possible in a practical test to match all dimensionless parameters. It is most important to

Dynamic similarity parameters

Nondimensional parameter*	Name	Physical effect
$\rho VL/\mu$	Reynolds number	Viscosity
V/c	Mach number	Compressibility
V^2/Lg	Froude number	Gravity
x_m/L	Knudsen number	Continuum
$\rho V^2 L/\sigma$	Weber number	Surface tension
$c_p\mu/\kappa$	Prandtl number	Heat conduction
$\beta TgL^3\rho^2/\mu^2$	Grashof number	Free convection

*The reference variables are defined as follows; L, length; μ, coefficient of viscosity; c, speed of sound; g, acceleration of gravity; σ, surface tension; κ, coefficient of thermal conductivity; β, coefficient of thermal expansion; T, temperature; x_m, mean free path of molecule; and c_p, specific heat at constant pressure.

match parameters that represent the dominant physical effects. Thus, correct simulation of viscous effects requires that Reynolds number be matched; Mach number may be ignored if compressibility effects are not important. In ship model tests, Froude number must be matched to duplicate wave patterns; the effect of Reynolds number on viscous drag may be predicted analytically. *See* DIMENSIONAL ANALYSIS; DIMENSIONLESS GROUPS; FLUID MECHANICS; FROUDE NUMBER; MACH NUMBER; MODEL THEORY; REYNOLDS NUMBER.

Alan T. McDonald

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Dynamical time

The independent variable in the theories and ephemerides of solar system bodies. Dynamical time scales are appropriate for coordinate systems in the solar system. *See* EPHEMERIS.

Types of time scales. There are different types of time scales in existence. Dynamical time scales are the appropriate coordinate time scales for the barycenter (center of mass) of the solar system, the geocenter (center of mass of the Earth), or planetary coordinate times. The different dynamical time scales are related by equations based on the gravitational potentials and the theory of relativity.

The most accurate time scales are based on physics, such as atomic time scales (with a second defined as the duration of 9,192,631,770 periods of radiation of a hyperfine transition of cesium-133), and International Atomic Time (TAI), an average atomic time coordinated by the Bureau International des Poids et Mesures in France. TAI is an atomic time for the Earth geiod, a surface of constant gravitational potential that is an extension of the surface of mean sea level of the Earth's oceans, and is the basis for many time scales. Universal Time is based on the rotation of the Earth and the motion of the Sun. Coordinated Universal Time (UTC) is based on TAI and differs from UT1, the observed Earth rotation time, by less that 0.9 s. The difference is kept less than 0.9 s by inserting leap seconds when necessary at the end of June or December. Civil time scales differ from UTC according to time zones mostly by hour increments. *See* ATOMIC TIME.

Terrestrial Time (TT) is a theoretically ideal proper time. A practical realization is TT = TAI + 32.184 s. The quantity ΔT = TT − UT1 is the difference between this ideal time scale and the rotation of the Earth. Before 1991, TT was known as Terrestrial Dynamical Time (TDT), a successor of prerelativistic Ephemeris Time (ET).

Geocentric Coordinate Time (TCG) is coordinate time for the geocenter. It is related to TT by Eq. (1),

$$\mathrm{TCG} = \mathrm{TT} + L_G \times (\mathrm{JD} - 2{,}443{,}144.5) \times 86{,}400\,\mathrm{s} \tag{1}$$

where JD is the Julian date, a widely accepted way of counting days consecutively, and $L_G \equiv 6.969290134 \times 10^{-10}$, a scale constant accounting for the Earth's gravitational and rotational potential affecting the rate of clocks operating at the geoid versus the geocenter. *See* CALENDAR; YEAR.

Barycentric Coordinate Time (TCB) is the coordinate time for the barycenter of the solar system. It differs from TCG by both secular and periodic terms. The complicated relationships between TT, TCG, and TCB are based on relativistic Lorentz transformations, although an approximation of (TCB − TCG) can be expressed as Eq. (2). Here TT_0 corresponds to JD 2,443,144.5 (1977 January 1, 0 h) TAI, the values of L_C and L_B are standard constants, $\mathbf{v}_e$ is the coordinate velocity of the Earth, $\mathbf{x}_e$ is the coordinate of the Earth's center of mass, and $\mathbf{x}$ is the coordinate of an event in barycentric reference coordinates. Periodic terms, denoted by $P(\mathrm{TT})$, have a maximum amplitude of around 1.6 m. *See* LORENTZ TRANSFORMATIONS; RELATIVITY.

$$(\mathrm{TCB} - \mathrm{TCG}) = \frac{L_C \times (\mathrm{TT} - \mathrm{TT}_0) + P(\mathrm{TT}) - P(\mathrm{TT}_0)}{(1 - L_B)} + \frac{\mathbf{v}_e \cdot (\mathbf{x} - \mathbf{x}_e)}{c^2} \tag{2}$$

Barycentric Dynamical Time (TDB) was originally intended to serve as an independent argument of barycentric ephemerides and equations of motion, and was defined by an International Astronomical Union 1976 resolution to differ from TT only by periodic terms. Later, it became clear that this condition cannot be rigorously satisfied. Operationally, each ephemeris realization defines its own version of TDB; the linear drift between TDB and TCB is usually chosen so that the rate of TDB matches TT as closely as possible for the time span covered by the particular ephemeris. TDB is sometimes designated as Barycentric Ephemeris Time ($\mathrm{T_{eph}}$) when used as the time scale of the Jet Propulsion Laboratory Development Ephemerides. When TDB is used, the units of length and mass are slightly different from the geocentric reference systems, and the astronomical unit

needs to be rescaled, depending upon the frame of reference and the accuracy required.

History. It was recognized that the rotation of the Earth was irregular. Thus, in 1952 the International Astronomical Union adopted Ephemeris Time, based on Simon Newcomb's theory of the Earth's motion as a uniform time scale. It became recognized that Ephemeris Time could be determined better from observations of the Moon than of the Sun. Also, in 1955 the more accurate atomic time became available, and it became the basis for Ephemeris Time. With improvements in accuracy, the relativistic distinctions between coordinate times had to be introduced. Hence, TDT and TDB were introduced in 1984. Difficulties with the names, definitions, and accuracies led to the introduction of TT in 1991 and of TCB and TCG in 2000 to replace TDT and TDB. *See* TIME.

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Dynamics

That part of classical mechanics which deals with the relation between the motions of material bodies and the forces acting upon them.

Basic concepts. Dynamics proceeds by adopting certain intuitively acceptable concepts which are associated with measurable quantities. These essential concepts and the measurable quantities used for their specification are as follows:

1. Space configuration refers to the positions and orientations of bodies in a reference frame adopted by the observer. It is expressed quantitatively by an arbitrarily chosen set of space coordinates, of which cartesian and polar coordinates are examples. All space coordinates rest on the notion of distance measurement. *See* FRAME OF REFERENCE.

2. Duration is expressed quantitatively by time measured by a clock or comparable mechanism. *See* ACCELERATION.

3. Motion refers to change of configuration with time and is expressed by time rates of coordinate change called velocities and time rates of velocity change called accelerations. The classical assumption that coordinates behave as analytic functions of time permits representation of velocities and accelerations as first and second derivatives, respectively, of the space coordinates with respect to time. *See* VELOCITY.

4. Inertia is an attribute of bodies implying their capacity to resist changes of motion. A body's inertia with respect to linear motion is denoted by its mass. *See* INERTIA.

5. Momentum is an attribute proportional to both the mass and velocity of a body. Momentum of linear motion is expressed as the product of mass and linear velocity. *See* MOMENTUM.

6. Force serves to designate the influence exercised upon the motion of a particular body by other bodies, not necessarily specified. A quantitative connection between the motion of a body and the force applied to it is expressed by Newton's second law of motion, which is discussed later. *See* FORCE.

Distance, time, and mass are commonly regarded as fundamental, all other dynamical quantities being definable in terms of them. *See* MASS; TIME.

Newton's second law. A primary objective of classical dynamics is the prediction of the behavior of bodies which are subject to known forces when only initial values of the coordinates and momenta are available. This is accomplished by use of a principle first recognized by Isaac Newton. Newton's statement of the principle was restricted to the linear motion of an idealized body called a mass particle, having negligible extension in space.

The basic dynamical law set forth by Newton and known as his second law states that the time rate of change of a particle's linear momentum is proportional to and in the direction of the force applied to the particle. This statement, although special in form, serves as a basis for more comprehensive statements of the principle which have since appeared.

Stated analytically, Newton's second law becomes the differential equation (1), in which m represents

$$\frac{d(mv)}{dt} = F \tag{1}$$

the particle's mass, v its velocity, F the applied force, and t the time. Equation (1) provides a definition of force and of its units if units of mass, distance, and time have previously been adopted. The classical assumption of constancy of mass permits Eq. (1) to be expressed as Eq. (2), where a represents the lin-

$$ma = F \tag{2}$$

ear acceleration. A particle in physical space requires three cartesian coordinates, x, y, and z, to specify its position. Its linear acceleration is a vector with three cartesian components, the second time derivatives of x, y, and z. Equation (2) therefore equates two vectors, requiring equality of their components expressed in detail by Eqs. (3). These are the newtonian

$$m\frac{d^2x}{dt^2} = F_x \tag{3a}$$

$$m\frac{d^2y}{dt^2} = F_y \tag{3b}$$

$$m\frac{d^2z}{dt^2} = F_z \tag{3c}$$

equations of motion of an unconstrained particle in space. If the three force components F_x, F_y, and F_z

are expressed functions of the coordinates and time, the dependence of each coordinate upon the time is implied and can in favorable cases be found as solutions of the equations of motion in the form of Eqs. (4). The primary objective of dynamics is

$$x = x(t) \quad y = y(t) \quad z = z(t) \tag{4}$$

achieved in the discovery of such functions.

One-dimensional particle motion. The motion of a particle which remains on the x axis, either because of constraints or initial conditions, is determined by Eq. (3*a*) alone, whose solution is simplified by the absence of y and z. Such one-dimensional dynamical problems provide an attractively simple introduction to the subject. Examples are the motion of a body falling vertically, subject to gravitational force, and linear harmonic motion.

Two-dimensional particle motion. The motion of a particle remaining in the plane of the x and y axes is determined by Eqs. (3*a*) and (3*b*), from which z is absent. Two-dimensional problems are reasonably tractable and include many of physical interest such as the motion of a projectile (exterior ballistics), and of a body attracted toward a central point, as in planetary motion. Solution of a two-dimensional problem is frequently simplified by change of variables which reduces it to a pair of one-dimensional problems.

Three-dimensional particle motion. All three equations of motion, Eqs. (3), apply to an unconstrained particle in space. Complete solutions are possible only when the functions expressing force components are relatively simple in character. Fortunately, many of the solvable cases correspond to important physical examples in which simplicity of the forces allows separation into one- and two-dimensional motions. Three-dimensional projectile motion without friction is an example.

Newton's third law. The behavior of systems composed of two or more interacting particles is treated by newtonian dynamics augmented by Newton's third law of motion which states that when two bodies interact, the forces they exert on one another are equal and oppositely directed. The important laws of momentum and energy conservation are derivable for such systems (the latter only for forces of special type) and useful in solution of problems. The equations of motion for systems of more than two interacting particles in space are mathematically intractable in the absence of geometrical constraints or special initial conditions, but assumptions approximating the physical situation permit solution of many problems of physical interest. The principles of particle dynamics are transferred to extended bodies by regarding them as systems of particles subject to specified mutual constraints and mutual forces. *See* BALLISTICS; GRAVITATION; HARMONIC MOTION; ORBITAL MOTION; RIGID-BODY DYNAMICS.

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Dynamometer

A device for measuring the torque, force, or power available from a rotating shaft. The shaft speed is measured with a tachometer, while the turning force or torque of the shaft is measured with a scale or by another method. Horsepower (hp) may be read from the instrumentation or calculated from Eq. (1),

$$\text{hp} = \frac{2\pi NT}{33,000} \tag{1}$$

where N is the shaft speed in rpm and T is the torque in pound-feet. Power may also be found from Eq. (2) or Eq. (3), where kW represents kilowatts.

$$\text{hp} = \text{torque (lbft)} \times \frac{\text{rpm (shaft speed)}}{5252} \tag{2}$$

$$\text{kW} = \text{torque (newton-meter)} \times \frac{\text{rpm (shaft speed)}}{9549} \tag{3}$$

See TACHOMETER.

Types. The two types are the transmission dynamometer and the absorption dynamometer. The transmission dynamometer transmits the force while measuring the elastic twist of the output shaft. An absorption dynamometer absorbs the power and dissipates it as heat by restraining the output shaft mechanically with a friction brake, hydraulically with a water brake, or electrically with an electromagnetic force. Since the restraining element tends to rotate with the output shaft, the force of the shaft can be determined by measuring the force required to arrest the rotation of the restraining element. Torque is then calculated by multiplying the force times the length of the lever arm, or the distance through which the force acts. The oldest type of absorption dynamometer is the prony brake, which mechanically applies a friction load to the output shaft by means of wood blocks, flexible bands, or other friction surface. *See* TORQUE.

One type of electric dynamometer consists of a direct-current (dc) machine with the stator cradle-mounted in antifriction bearings. The rotor is connected to the shaft of the machine under test. The field current is introduced through flexible leads. The stator is constrained from rotating by a radial arm of known length to which is attached a scale for measuring the force required to prevent rotation. The torque of the connected machine is found from the product of the lever arm length and the scale reading, after correcting the scale reading by the amount of the zero torque reading.

Uses. The most common use of the dynamometer is in determining the power of an electric motor or engine of a car, truck, or other vehicle. A dynamometer that connects to the engine crankshaft is an engine dynamometer. One that has rollers turned by the vehicle drive wheels is a chassis dynamometer;

this type is widely used in the automotive industry for mileage accumulation, emissions, fuel economy, and performance testing of cars and trucks. For fuel-economy testing, dynamometer inertia simulation may be provided with either mechanical-flywheel equivalent weights or electrical load.

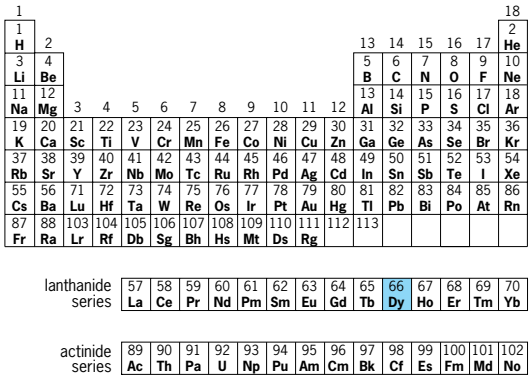
If the machine under test is a motor, the dynamometer will act as a generator. Dynamometer output is then absorbed by a loading resistance or by feeding it into a direct-current line. The amount of output is adjusted by changing the loading resistance or the field excitation. If the machine under test is a generator or mechanical load, the dynamometer will act as a motor. Speed is adjusted by changing the armature voltage or the field excitation. *See* DIRECT-CURRENT GENERATOR; DIRECT-CURRENT MOTOR; ELECTRIC ROTATING MACHINERY; GENERATOR; MOTOR.

Donald L. Anglin

Dysprosium

A metallic rare-earth element, Dy, atomic number 66 and atomic weight 162.50. The naturally occurring element is composed of seven stable isotopes. Dysprosium forms a white oxide, Dy_2O_3, which dissolves in acid to give a yellowish-green solution. *See* PERIODIC TABLE; RARE-EARTH ELEMENTS.

The metal is attacked readily by air at high temperatures, but at room temperatures, in massive blocks, it is fairly stable in the atmosphere and remains shiny for long periods of time. Dysprosium is paramagnetic, but as the temperature is lowered, it becomes antiferromagnetic at the Néel point (178 K or −139°F) and ferromagnetic at the Curie point (85 K or −306.4°F). At very low temperatures, the metal shows strong anisotropic magnetic properties.



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e (mathematics)

The number e usually defined as the limit approached by the expression

$$\left(1+\frac{1}{n}\right)^n$$

as n approaches infinity. If the given expression is expanded by the binomial theorem, and if one uses the theorem that the limit of the quotient of two polynomials of equal degree as the variable tends to infinity is equal to the ratio of the coefficients of highest degree, one obtains the expansion in relation (1).

$$e = 1 + \frac{1}{1} + \frac{1}{1\cdot 2} + \frac{1}{1\cdot 2\cdot 3} + \cdots + \frac{1}{1\cdot 2\cdot 3\cdot n} + \cdots \quad (1)$$

Clearly e is larger than 2; it may be easily shown that it is smaller than 3. It can also be shown by elementary methods that e is irrational; that is, it cannot be represented as the quotient of two integers. Furthermore, e is transcendental; it does not satisfy any algebraic equation with integral coefficients. The transcendentality of e was proved by the French mathematician C. Hermite in 1873; the proof constitutes an important milestone in the history of mathematics.

By the method outlined above, it may be shown that the limit of

$$\left(1+\frac{x}{n}\right)^n$$

as n tends to infinity is e^x, and moreover that relation (2) holds. The function e^x is of great importance in mathematical analysis and is encountered in numerous problems in applied mathematics. For this reason it has been extensively tabulated. The most recent tables of exponentials for both positive and negative values of x were computed by the Computation Laboratory of the National Bureau of Standards. One of the most important formulas in mathematics involving e is Euler's formula $e^{i\theta} = \cos\theta + i\sin\theta$. This is readily obtained from relation (2) by replacing x by $i\theta$. The real part of the expansion is recognized as the expansion of $\cos\theta$, while the coefficient of i is recognized as the expansion of $\sin\theta$. An immediate consequence of Euler's formula is de Moivre's formula $(\cos\theta + i\sin\theta)^m = \cos m\theta + i\sin m\theta$. If the first member of the latter formula is expanded by the binomial formula, and the real and imaginary parts of both members are equated, one obtains two important formulas which permit the evaluation of $\cos m\theta$ and $\sin m\theta$ in terms of $\cos\theta$ and $\sin\theta$. *See* BINOMIAL THEOREM; CALCULUS; LOGARITHM.

$$e^x = 1 + \frac{x}{1} + \frac{x^2}{1\cdot 2} + \cdots + \frac{x^n}{1\cdot 2\cdot 3\cdots n} + \cdots \quad (2)$$

Arnold N. Lowan; Salomon Bochner

Ear (vertebrate)

The organ which sends information about sound to the brain, constituting the sense of hearing, as well as vestibular information about the orientation of the head in space. The vertebrate ear is generally divided into three regions that have discrete functions: The inner ear is found in all vertebrates, and it subsumes both hearing and balance (functions). The external ear and the middle ear, not found in all vertebrates, enhance hearing. *See* HEARING (VERTEBRATE).

Ear structure. The inner ear is embedded in the ear (or otic) capsule and has a common embryological development in all vertebrate groups. As the presumptive brain forms in the early embryo, an ectodermal thickening, the auditory placode, appears on each side of the head surface. This balls up, sinks inward, and is then pinched off to form a hollow sac of epithelial tissue (called an otocyst) within what will become the otic bones just lateral to the posterior part of the hindbrain. The embryonic cells of the otocyst will differentiate to form all the epithelial portions of the inner ear. Surrounding mesenchymal

tissue forms the connective tissue of the inner ear. As differentiation proceeds, tissue forms and develops the nerve (cochlear-vestibular, or eighth cranial, nerve) that carries information from the inner ear to the brain. In this manner, the otocyst and immediately surrounding embryonic tissues form all parts of the inner ear, auditory and vestibular.

The middle ear and external ear are not found in the fishes (including jawless fishes, sharks, and bony fishes). All tetrapods (amphibians, reptiles, birds, and mammals) have a middle ear with a tympanic membrane. Reptiles, birds, and mammals also have an external auditory meatus (or canal) which extends from the tympanic membrane to the external surface of the head. Mammals generally have an external structure, the pinna, that helps "collect" and carry the sound to the ear canal and then to the tympanum. The major difference in the middle ear among tetrapods is that it has a single ear bone, or ossicle (often called the columella or stapes), in amphibians, reptiles, and birds, while mammals have three middle-ear bones (malleus, incus, and stapes).

In comparing the inner ears of different vertebrates, from sharks and fishes to mammals, the major structural differences are associated with the auditory part of the ear. With few exceptions (notably in some fishes that detect very low or very high frequency sounds), the vestibular portion of the inner ear is developmentally, structurally, and functionally nearly the same in all vertebrates.

Sensory hair cells. The basic sensory unit in the inner ear is the sensory hair cell (**Fig. 1**). These specialized cells are morphologically similar in all of the epithelial structures of the ear in all vertebrates (and in the lateral line of fishes and amphibians), but they may have either auditory or vestibular functions depending upon the associated superstructure. The superstructure serves to facilitate the transmission of vibrations from the environment to the hair cells. For the vestibular apparatus, the superstructure blocks external vibratory energy, but sensitizes the sensory hair cells to the pull of gravity and to acceleratory and deceleratory movements of the head. *See* LATERAL LINE SYSTEM.

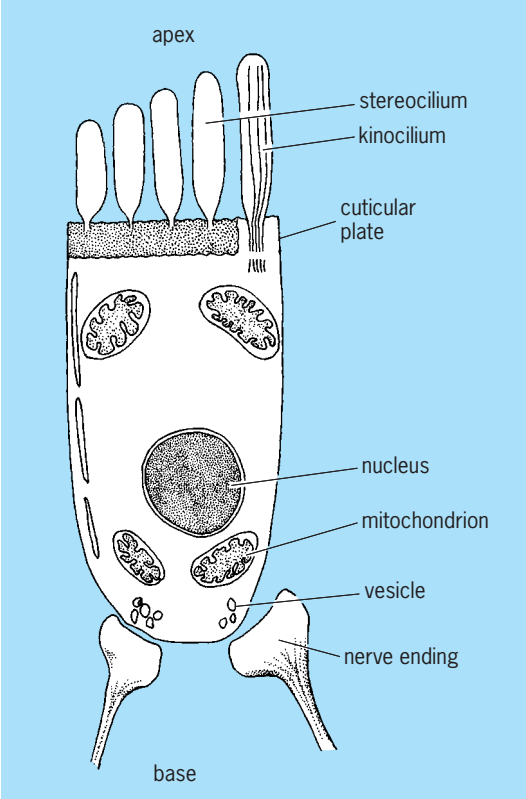

Fig. 1. Generalized diagram of a sensory hair cell found in the ear of all vertebrates.

The sensory hair cell is a columnar, polarized structure from whose apex extend thin cilia that resemble hairs. Each hair cell has many such cilia, making up a ciliary bundle which bends in response to motional energy. The cilia in each bundle include many stereocilia and a single, eccentrically positioned kinocilium. The stereocilia are a homogeneous structure within a simple membrane covering, while the kinocilium contains one pair of microfibrils in the center and nine pairs around the periphery. The stereocilia are arranged in an orderly pattern, and the kinocilium is located at one end of the stereociliary bundle. The cilia extend into an extracellular fluid-filled space, with their tips embedded in a gelatinous membrane. *See* CILIA AND FLAGELLA.

The body of the sensory hair cell lies below the ciliary bundle. Just below the cell membrane, at the apex of the cell, lies a relatively homogeneous electron-dense region, the cuticular plate, from which the stereocilia emerge (Fig. 1). The kinocilium also arises from the cellular apex, but from a pair of basal bodies that are located to the side of the cuticular plate. Immediately below the cuticular plate lies a group of mitochondria, the energy-producing organelles found in all cells. Nearer the basal part of the cell is the nucleus, and below that, another group of mitochondria and a group of small, vesiclelike structures similar to the synaptic vesicles of nerve endings. These structures contain neurotransmitters that are similar to those found in other neuronlike cells in the nervous system. In the sensory hair cell, the vesicles release transmitters into the synaptic space between the cell and the neurons of the eighth cranial nerve that innervate the cells. These transmitters excite the neuron, which then sends on information to the brain. *See* MITOCHONDRIA; NEUROBIOLOGY; SYNAPTIC TRANSMISSION.

The sensory hair cell is the detector of motion, either produced by compression and rarefaction of molecules due to sound waves, or imparted by movement of the head against gravity. This motion produces bending of the ciliary bundles, and this in turn results in a change in configuration of the membrane overlying the stereocilia and opening of channels in the membrane. It is generally thought that these channels admit calcium into the cell, and this in turn interacts with other components of the cell. Ultimately, the energy generated by these interactions causes release of neurotransmitter at the base of the cell, and results in stimulation of afferent neurons which contact the cell.

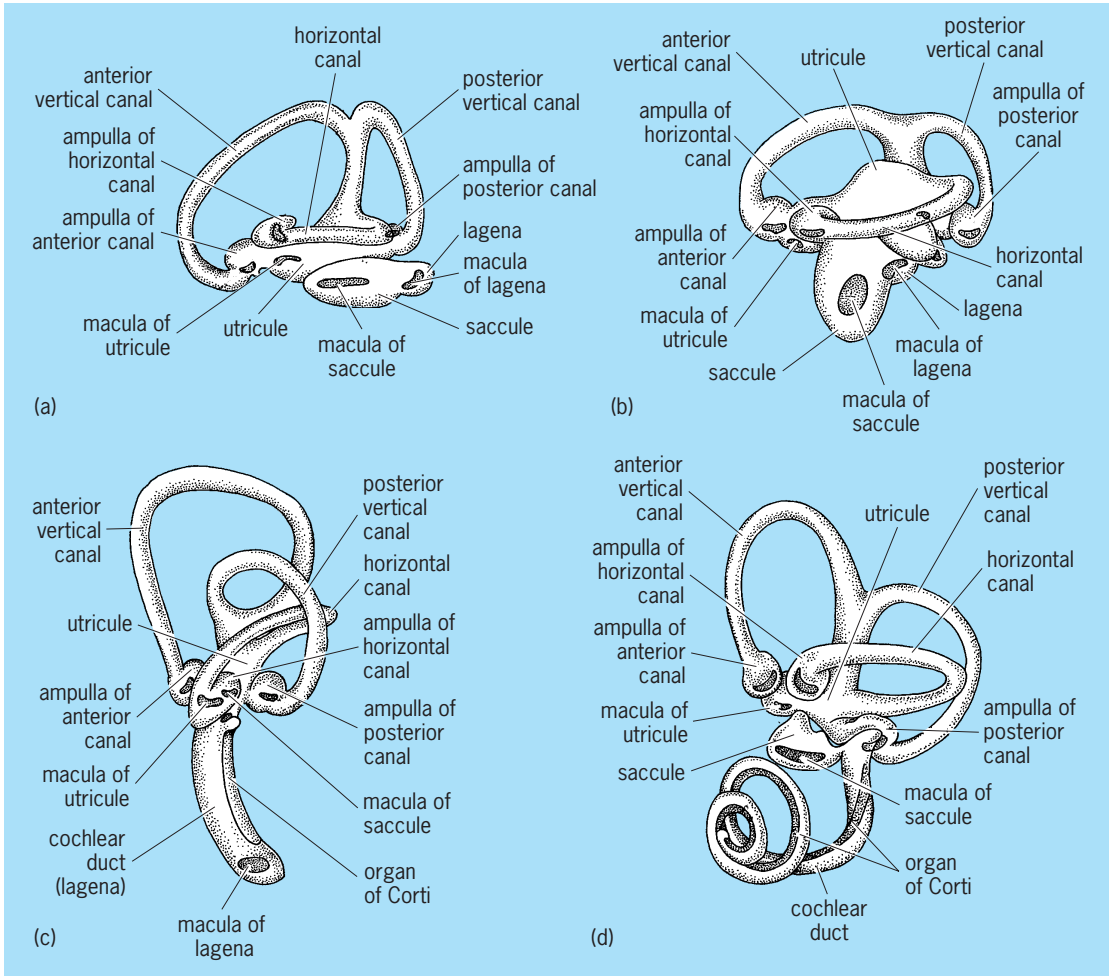

Fig. 2. Left-ear external view of the membranous labyrinths of (*a*) teleost, (*b*) frog, (*c*) bird, and (*d*) mammal. (*After A. S. Romer, The Vertebrate Body, 3d ed., Saunders, 1962*)

Regeneration of sensory hair cells. Once it was widely believed that all sensory hair cells found in a vertebrate are present at birth, and that the only change in number of these cells is a decrease due to cell death as a result of aging or damage produced by chemicals or loud sounds. However, recent findings have shown that both ideas are incorrect, at least in several vertebrate groups. It is now known that elasmobranchs (sharks and rays) and fishes continue to produce additional sensory hair cells throughout their lives (as long as they grow). As a result, the number of sensory hair cells in a large shark or fish is much greater than in a younger animal. Amphibians also add sensory hair cells after birth, but in smaller numbers than in fishes, and there is some evidence that such addition occurs in the vestibular part of the ear in birds. However, mammals do not add sensory hair cells after birth.

It has also been shown in fishes and birds that hair cells will regenerate after trauma induced by certain drugs (such as streptomycin and related drugs) and loud sounds. Both stimuli kill sensory hair cells, but after several days or weeks these cells regenerate, and it now appears that full function returns to the ear, at least in birds. Again, there is no evidence to indicate that destroyed hair cells are replaced in mammals.

Fishes. While the basic structure of the inner ear and of the sensory hair cells is the same in all vertebrates, a number of interesting differences reflect the evolution of the ear and the ways in which different animals detect, process, and use sound (**Fig. 2**).

In elasmobranchs and bony fishes the inner ear is located in the brain (cranial) cavity somewhat behind the eye (**Fig. 3**). The inner ear has several regions, including three semicircular canals and otolith organs. Other than the very primitive jawless fishes (agnathans: lamprey and hagfish) which have one or two semicircular canals, all other vertebrates have three canals. All fishes, amphibians, reptiles, and birds have three otolith organs—the saccule, utricle, and lagena—while mammals do not have the lagena. It has been suggested by some investigators that the lagena gave rise to the cochlea, which is the main auditory organ of mammals.

Each of the three semicircular canals lies on one plane of space. The horizontal canal is in the horizontal plane relative to the animal's normal head position; the anterior and posterior canals are vertical

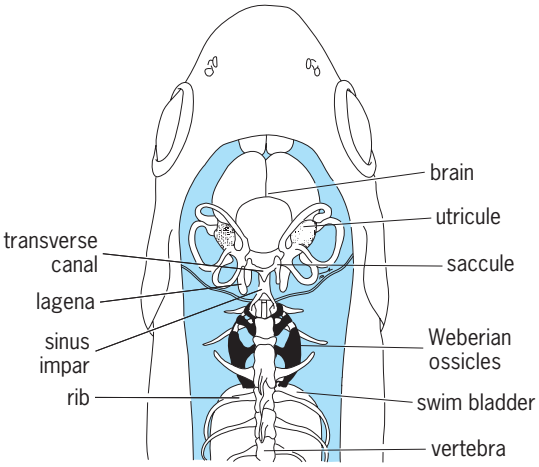

Fig. 3. Dorsal view of the anterior part of the head and body of an otophysan fish showing the ear and its position relative to the brain. Also shown are the swim bladder and the Weberian ossicles. (*From A. N. Popper and S. Coombs, Auditory Mechanisms in Teleost Fishes, Amer. Sci., 68:429–440, 1980, redrawn from K. von Frisch, Über den Gehörsinnder Fische, Biol. Rev., 11:210–246, 1936*)

and at right angles to each other as well as to the horizontal canal (Fig. 2*a*). While the names of the canals differ in humans, due to the upright position of the body, the canals in humans are identical in form and function to those in all other vertebrates.

At one end of each endolymph-filled semicircular canal is a widened area, the ampulla, which has a sensory area called the crista (or crista ampullaris). The crista contains large numbers of sensory hair cells, as well as other cells which provide support for the hair cells. At the base of the hair cells are nerve endings from the vestibular branch of the eighth cranial nerve.

The long kinocilia and stereocilia of the hair cells of the crista extend into a gelatinous structure, the cupula. The cupula blocks the semicircular canal at this point and prevents the endolymph from free movement into the ampulla. Upon acceleration or deceleration of the animal's head, movement of the fluid in the semicircular canal of the corresponding plane lags slightly behind the movement of the membranous canal and of the head itself. When the fluid catches up, it puts pressure upon the cupula, bending the hair cells and initiating nerve impulses. Since more than one semicircular canal can react at a time, depending upon the precise movement of the head, the brain receives very precise information about head movement.

Each of the otolith organs also has a sensory area, called a macula, that contains hair cells and supporting cells. The cilia of the otolithic organs are not nearly as long as those found in the ampullae of the semicircular canals. They are, however, embedded in a thin gelatinous membrane that also contains very dense calcium carbonate crystals. In elasmobranchs, as in primitive fishes and all tetrapods, these crystals are called otoconia. In most bony fishes the crystals are fused into a single mass in each otolith organ called the otolith. The otoliths generally have very species-specific shapes and grow as the fish grows by adding daily and yearly rings. By proper preparation of the otoliths, and particularly those of the saccule, it is possible to tell the age of a fish, as well as the quality of diet and other physiological characteristics about the animal. In some ways, this is analogous to dating a tree by the number of growth rings.

Hearing. Fishes are able to detect a wide range of sound using their inner ear. Most fishes detect sounds from 30 to 800 or 1000 Hz, with best hearing from 200 to 500 Hz. However, some fishes, called hearing specialists (such as goldfish and catfish), have evolved special mechanisms to enhance hearing to 3000 or 4000 Hz. A few species in the herring family can detect sounds to almost 200,000 Hz, and it is thought that this capability has evolved to enable these species to detect echolocating signals of their major predators, dolphins, and avoid being eaten. Significantly, fish not only can detect the presence of sound but also are able to discriminate between sounds, determine the direction of a sound source, and detect sounds in the presence of noise, as can mammals.

Many species of fish also produce sounds and use sound for communication, as in reproduction and territorial behavior. In these species, hearing is needed for communication with other animals. Other species, such as elasmobranchs, detect prey by the sounds they make. However, many species of fish hear well and yet are not known to produce sounds or use sounds to find food (such as the goldfish). It is thought that hearing evolved so the animals can gain general information about their environment from the noises that are constantly found in the sea. In many ways, this is similar to the kind of information that a person gets from sound upon entering a dark room. Thus, it is likely that the use of sound for communication came after the sense of hearing evolved.

Tetrapods detect sounds that impinge on the tympanic membrane and then are carried by the middle-ear bones to the inner ear, where the sounds set the fluids of the ear into motion and thus stimulate the sensory hair cells. In fishes, however, this kind of pathway is not needed since sound is already traveling through water. Indeed, since the fish's body is the same density as that of water, sound would travel right through the fish were it not for the otoconia or otoliths. Since these structures are much denser than the fish's body and the water, they stay still while the fish's body and the attached sensory hair cells move with the sound field. Since the stereocilia are attached to both the top of the hair cells and to the otoconia or otolith, they are bent as their base moves with the macula and their tops stand still with the otoliths. This bending sends signals to the nerves and then to the brain, indicating the presence of a sound.

Swim bladder. Some species of fish, the hearing specialists, use a secondary structure, the swim bladder, to enhance hearing capabilities. The swim bladder is a bubble of gas found in the abdominal cavity of most

bony fishes (but not elasmobranchs), and it is used primarily for buoyancy control, though it may also be used in sound production in some species. Since the swim bladder is filled with gas, its density is different from that of the rest of the fish, and in a sound field the walls of the swim bladder are set into vibration and act as a small sound source to send sounds to the ear. This method enhances hearing only in the hearing specialist species, where the front end of the swim bladder comes close to the saccule or utricle, or there is some bony connection that serves in a manner analogous to the tetrapod middle-ear bones. The best-known example is found in the group of fishes called the Otophysi (catfish, goldfish, carp), where a series of bones, the Weberian ossicles, transmit the movement of the swim bladder walls to the saccule of the inner ear (Fig. 3). These hearing specialists can hear a wider range of frequencies than nonspecialists, and their hearing sensitivity at each frequency is also better. *See* SWIM BLADDER.

Investigators who first studied fish hearing in the early part of the twentieth century thought that the saccule and lagena were the auditory parts of the ear in fishes (pars inferior) while the utricle served vestibular functions along with the semicircular canals (pars superior). However, more recent studies strongly suggest that all three otolith organs in fishes are involved in both the vestibular and the hearing senses.

Tetrapods. Many structural and functional features of the fish inner ear are also found in the tetrapod ear. The inner ear of tetrapods is embedded in the otic bones of the skull, with the membranous labyrinth attached to the bony labyrinth by connective tissue but suspended in perilymphatic fluid. There are three semicircular canals, with cristae, and, except in mammals which do not have a lagena, the three otolithic organs (Fig. 2*b*, *c*, *d*). In their morphology and physiology the vestibular parts of fish and tetrapod ears are nearly the same. For the most part, the tetrapod otolithic organs function only as vestibular organs rather than playing an auditory role as they do in fishes.

Adaptations for aerial habitat. Audition in land vertebrates poses special problems not met by fishes. These problems arise from the aerial (as opposed to the aquatic) environment and from the different specific gravities of environment and organism.

The auditory inner ear is in a fluid environment and must receive sound vibrations from an aerial environment (unlike in fishes, where the sound is carried in basically the same fluid as found in the inner ear). Because air is much less dense and more elastic than water, almost all sound energy in air is reflected from a fluid surface rather than being transmitted to the fluid. In physical terms, this is called an impedance mismatch between air and fluid. In order for the sound energy in air to be transmitted to fluids, a mechanism must be present to match the impedance of air with that of the fluids. A combination of increasing the sound pressure and decreasing the particle velocity at the interface between air and fluid can match their impedances. This is the primary function of the tetrapod middle-ear apparatus. *See* SOUND.

With the emergence of vertebrates on land, gilled respiration was lost. The gill slits became obliterated except for the most anterior, which, highly modified, became the middle-ear cavity, retaining its connection to the mouth cavity in the structure called the eustachian tube. The cavity does not quite open to the outside, but is covered by a very thin layer of epithelium, called the tympanic membrane.

Amphibians. The tympanic membrane of frogs and toads is located on the lateral surface of the head. Attached to its inner aspect is a small rodlike bone, the stapes, or columella, which runs through the air space of the middle-ear cavity and plugs a small hole, the oval window, beyond which are the inner-ear fluids. The frog's tympanic membrane collects sound energy and transmits it through the columella to the inner-ear fluids. The surface area of the tympanic membrane is quite large, while the surface area of the columella footplate, where it meets the inner ear fluids, is much smaller. As a consequence, although the amount of energy collected by the tympanic membrane is the same as that received by the inner-ear fluids, the energy has been concentrated into a much smaller area by the time it reaches the fluids. Thus, the force per unit area (sound pressure) at the inner ear is much greater than it is at the tympanic membrane. This is the impedance-matching mechanism of the middle ear of amphibians. *See* AMPHIBIA.

Sensory areas. In the lagenar portion of the amphibian's membranous labyrinth are two areas of hair cells, the amphibian and basilar papillae, that are found in no other vertebrate group. The basilar papilla lies on the posterior wall of the saccule between the oval window and the round window, another membrane-covered opening between middle ear and inner ear. The ciliary bundles on the sensory hair cells of the basilar papilla reach into the endolymph and are embedded in the gelatinous tectorium, or tectorial membrane. This tectorium is thought to correspond to the structure of the same name in the ears of amniotes (tetrapods other than amphibians), but differs in that it not only receives cilia from the hair cells but also sends down strands of its own between the hair cells to attach to the supporting cells.

The amphibian papilla has a structure similar, although not identical, to the basilar papilla. Movements of the inner-ear fluids cause bending of cilia of the hair cells in both papillae, but they respond differentially to different frequencies. The precise frequency range that each responds to differs in different species, and it is related to the frequencies of the sounds used by a particular species for communication.

Sound reception. Vibratory energy enters the inner ear at the oval window, passes through the basilar and amphibian papillae causing them to vibrate, and then dissipates at the round window. Physiological experiments on frogs and toads have shown that the basilar and amphibian papillae are the primary organs receiving sound and transducing its physical

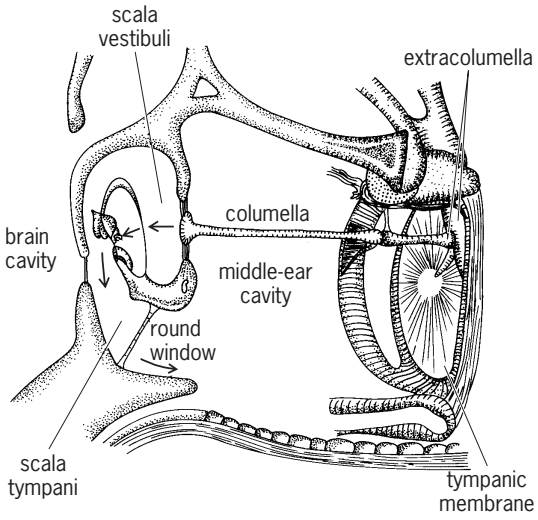

Fig. 4. Diagrammatic view of the middle and inner ears of a lizard. (*After D. Webster and M. Webster, Comparative Vertebrate Morphology, Academic Press, 1974*)

vibrations into nerve impulses. However, some data suggest that the saccule may also serve an auditory function in frogs and toads, particularly for the detection of very low frequency vibrations, and even seismic motion.

Birds and reptiles. In most reptiles and birds the tympanic membrane lies not on the surface of the head but internally, at the end of the tube called the external auditory meatus. A middle-ear cavity (with its eustachian tube to the mouth) lies medial to the tympanic membrane. A single ossicle, the columella, crosses this cavity from the tympanic membrane to the oval window at the inner ear. As in amphibians, the surface area of the tympanic membrane is much greater than that of the footplate of the columella. As a result, the force per unit area is thus greatly increased, and aerial vibrations cause fluid vibrations in the inner ear (**Fig. 4**). *See* AVES; REPTILIA.

Auditory end organ structure. While both birds and reptiles have saccule, utricle, and lagena, as well as semicircular canals, they also have a newly evolved end organ, the basilar papilla, which is the part of the ear used for hearing in both groups of animals. (The avian and reptilian basilar papilla is thought to be a totally different structure, in terms of evolution and embryonic origin, than the basilar papilla found in amphibians). The lagena lies at the end of the elongate basilar papilla (Fig. 2*c*), and so it has been suggested by some investigators that the basilar papilla is derived from the lagena. However, there are few data to support this suggestion. The basilar papilla in birds and reptiles is often also called the cochlea, and there is some evidence to suggest that this end organ is directly related to the mammalian cochlea.

The basilar papilla in reptiles is generally somewhat shorter than that found in birds, and there is considerable variation in the specific structure of this end organ in different species. In particular, the basilar papilla in lizards shows wide variation. However, very little is known about the functional significance of this variation, and in fact, very little is known about hearing capabilities of reptiles since they are hard to test in behavioral experiments. At the same time, since the basilar papilla is so well developed in many species, it is likely that these animals are capable of detecting and discriminating a wide range of sounds. While there is less variability in the basilar papillae of birds, it is known that many of these species hear very well (some species to over 7000 Hz), and sound is a very important part of the avian communication system.

The basilar papilla contains sensory hair cells that are supported by sustentacular cells and arranged in rows, and each has a kinocilium and several stereocilia. In birds, the basilar papilla sensory hair cells are differentiated into short and tall hair cells, which may have different functions in hearing.

Efficiency of transduction. Above the basilar papilla the cilia of the hair cells are embedded in a gelatinous rectorial membrane whose precise shape differs markedly in different species. It is attached not to the supporting cells, as in amphibians, but to the spiral limbus, a ridge of epithelial cells firmly adhering to the bony labyrinth on one side of the basilar papilla. Vibrations in the inner ear cause the basilar membrane and its papilla to move, while the tectorial membrane, anchored on bone, does not move. The hair cells extending from the papilla and embedded in the tectorial membrane are thus bent, and more efficiently so than are those of amphibians, in which the membrane, the hair cells, and the supporting cells are attached only to each other.

Mammals. The mammalian ear consists of three parts: the external ear which receives the sound waves; the middle ear which transmits the vibrations by a series of three small bones; and the inner, or internal, ear, a complex bony chamber placed deep in the skull (**Fig. 5**). The external auditory meatus plus the newly evolved pinna, a cartilaginous structure

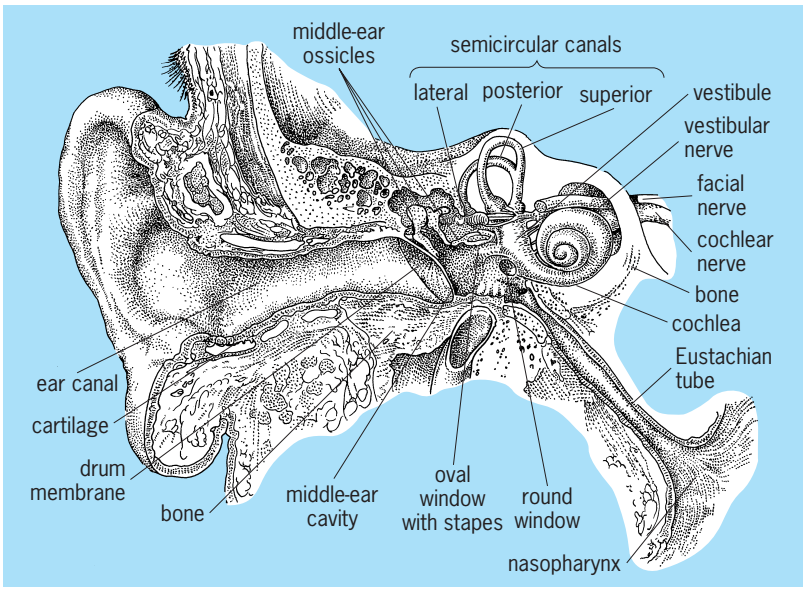

Fig. 5. Schematic drawing of the human ear. (*After M. Brödel, Three Unpublished Drawings of the Anatomy of the Human Ear, Saunders*)

projecting from the ear, compose the external ear. The shape and size of the pinna vary greatly. In elephants it is a huge, flapping structure, while in whales and moles it is completely absent. The size of the pinna in other species falls between the two extremes. The auditory function of the pinna varies widely in different species. In some species the pinna is moved in the direction of a sound source and helps the animal focus sound to the external auditory meatus and then down the ear canal. In other species, such as humans, the pinna may have a lesser function, but even in humans the pinna helps to discriminate between sounds coming from the front and back of the head so that the person can better tell the direction of a sound source. *See* MAMMALIA.

Middle ear. As in other tetrapods, the first gill slit is modified as a middle-ear cavity, communicating with the pharynx by way of the eustachian tube (**Fig. 6**). In other tetrapods this tube is permanently open, while in mammals it is usually closed. In order to equalize the pressure between the inside of the middle-ear cavity and the environment, the tube must be opened by yawning or swallowing. This is the "popping" in the ears which is experienced when one drives rapidly up a mountain or rides in an elevator.

Instead of the single columella of other tetrapods, the mammalian middle ear has three bones, closely articulated with one another. The innermost is the stapes, which fits into the oval window of the inner ear and is homologous with the columella. Attached to the tympanic membrane is the malleus, and lying between the malleus and stapes is the incus. These outer two bones are derived from the quadrate and articular bones, which in other tetrapods and most fishes are part of the jaw joint.

The geometry and arrangement of these three ossicles constitute a lever system which produces both an amplification of pressure and a decrease in particle velocity. However, the amplification of force per unit area, from large tympanic membrane to small stapes footplate, also occurs here. The adaptive value of three ossicles rather than one appears to be an added flexibility for hearing loud as well as soft sounds without endangering the delicate inner ear. In spite of having additional bones, the mammalian middle ear functions basically as do those of amphibians, reptiles, and birds in transforming aerial vibrations into fluid vibrations within the inner ear.

Inner ear. In the mammalian inner ear the vestibular apparatus is much like that of other tetrapods. The auditory portion, however, is considerably different (**Fig. 7**). It is elongated and coiled into a snail shape. This structure is called the cochlea. The epithelium of the basilar papilla is more differentiated in mammals than in other tetrapods, and it is called the organ of Corti. The number of turns in the cochlea varies from less than one in the generalized monotremes, to two and a half in humans, to five in the guinea pig. At the base of the cochlea is the oval window, which carries sound energy into the inner ear, and the round window, where this energy is dissipated after traveling in the cochlea.

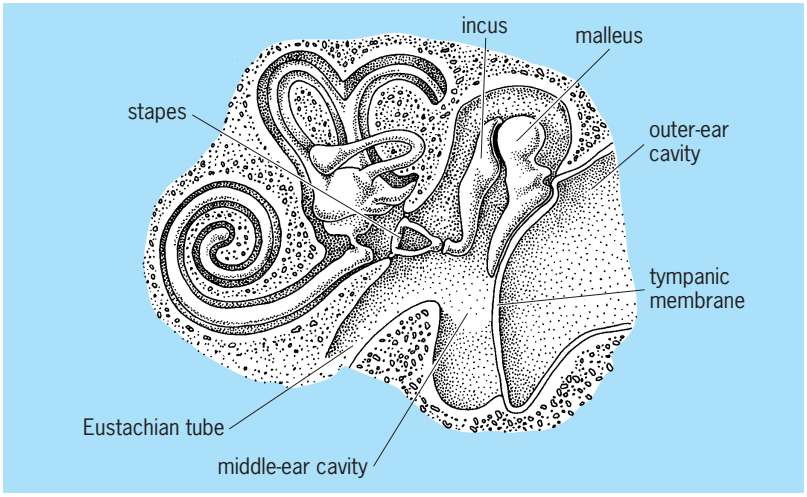

Fig. 6. Otic region of the heah of a mammal (showing the ear region only). (*A. S. Romer, The Vertebrate Body, 3d ed., Saunders, 1962*)

Running the length of the coiled cochlea are three channels; the uppermost, the scala vestibule, and the lowest, the scala tympani, communicate with each other at the helicotrema, an opening at the apex of the cochlea. These channels are filled with perilymph. In the center is the scala media, or cochlear duct. The cochlear duct is filled with endolymph, and it is separated from the scala vestibule above by the thin Reissner's membrane and from the scala tympani below by the basilar membrane.

The basilar membrane is suspended on both sides by ligaments or bone. The basilar membrane varies regularly in width, being narrow at the base (where it is most responsive to high frequencies) and wide at the apex (where it is most responsive to low frequencies).

Resting upon the basilar membrane is the organ of Corti. The organ of Corti contains several cell types in addition to the auditory hair cells. The hair cells are of two types, one type on either side of the pillar cells. The hair cells lying on the internal side of the pillar cells are called the inner hair cells, and those lying on the external side are called the outer hair cells. Morphologically, the inner hair cells are flask-shaped, whereas the outer hair cells are cylindrical. The external cell membrane of the outer hair cells is also lined by molecules which appear to act as molecular motors which can contract and cause a change in the length of the sensory cells. These motors are not found in inner hair cells, or the vestibular hair cells of mammals, or any hair cells in other vertebrates.

In both inner and outer hair cells of many mammals, including humans, the kinocilium is completely missing after birth of the animal, and only stereocilia are present. Typically, the single row of inner hair cells and three rows of outer hair cells run the length of the basilar membrane. Outside the hair cells and supporting cells are lines of border cells.

There may be up to 20,000 sensory hair cells in a cochlea of a normal young human, although the number of hair cells declines with age as a result of normal cell death, damage due to some medications,

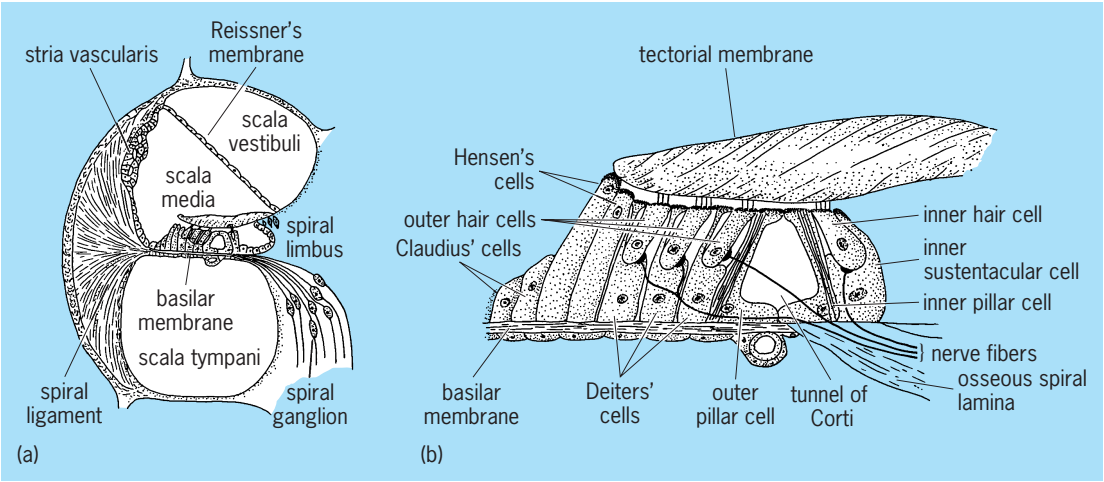

Fig. 7. Inner ear of a mammal. (*a*) Cross section of a single turn of the cochlea. (*b*) Organ of Corti. (*After D. Webster and M. Webster, Comparitive Vertebrate Morphology, Academic Press, 1974*)

and trauma caused by loud sounds. Indeed, hearing abilities of humans (and presumably other mammals) decline with age. A healthy teenager may hear sounds from below 20 Hz to upward of 20,000 Hz, while an adult 40 or 50 years old may hear sounds only to 14,000 Hz (or even less). This loss of hearing is associated with death of sensory hair cells.

Lining the external, or outer, wall of the scala media is a highly vascularized portion known as the stria vascularis. This area is metabolically very active and supplies the total blood and nutrients for the organ of Corti. Along the inner wall of the scala media (which is attached to the bone of the modiolus) is the area known as the spiral limbus. This is attached firmly to the gelatinous rectorial membrane which extends over the organ of Corti. In turn, the stereocilia of the sensory hair cells of the organ of Corti are embedded in the tectorial membrane. As in birds and reptiles, the tectorial membrane is held steady while the basilar membrane vibrates; the embedded hairs are bent, and transduction of physical energy into nerve impulses occurs (Fig. 5).

Sounds entering the mammalian inner ear at the oval window travel along the basilar membrane from basal to apical ends, causing vibrations of the membrane. The largest motion of the basilar membrane to a sound depends on the frequency components of the sound. Different frequencies maximally excite different regions of the basilar membrane, with high frequencies causing maximum motion of the basal end (the end near the oval window) and low frequencies maximally exciting regions closer to the apical end (near the helicotrema). This produces a "place mechanism" of hearing. When complex sounds, such as speech, enter the ear, different regions of the basilar papilla respond to different frequencies making up the sound. In effect, the basilar membrane is capable of breaking down a sound into its component frequencies.

The basis for responding to different frequencies at different regions of the basilar membrane is the stiffness of the membrane itself. The stiffness is a function of many factors, including the width and thickness of the membrane which varies along the length of the cochlea. The response of the different regions of the organ of Corti to specific frequencies is also thought to be enhanced by the sensory hair cells themselves. Whereas early investigations suggested that both inner and outer hair cells were involved in detection of sound per se, recent evidence suggests that the inner hair cells have the major role in hearing, while the outer hair cells modify the function of the ear and help to enhance the sensitivity of the inner hair cells.

Each of the outer hair cells is innervated by nerves which bring signals to the cells from the brain, while the inner hair cells have nerves which carry signals from the cells to the brain. The signals to the outer hair cells appear to tell the cells to contract or lengthen, in response to the nature of the sound being detected, and this change in length affects the stiffness of the membrane, and thus the sensitivity of the inner hair cells. *See* HEARING (HUMAN).

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Ear protectors

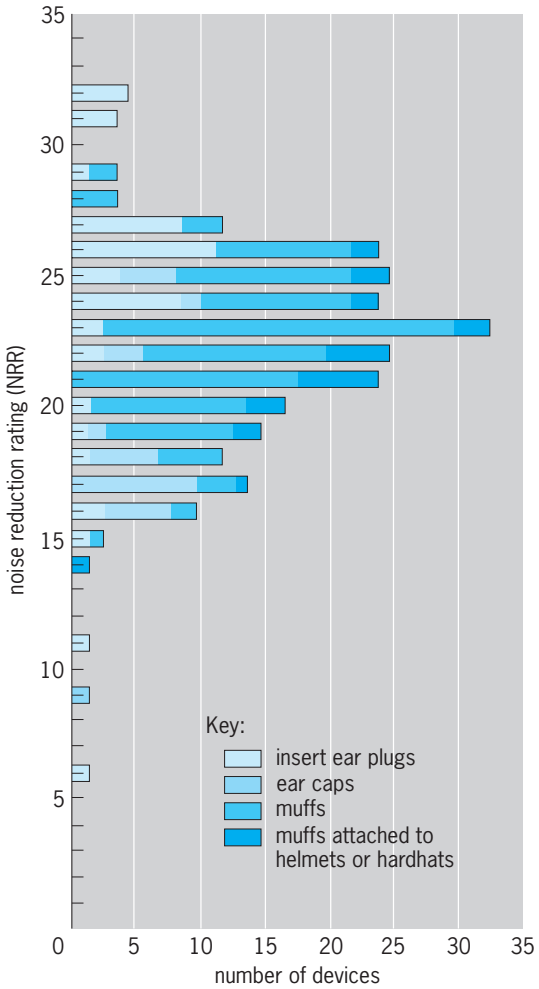
Objects or devices which insert into the ear canal or cover the outer ear and which have a variety of applications. They may protect the ears from

exposure to caustic materials, extremely cold temperatures, or excessive noise; serve as a barrier against solid particles collecting on the surface of the outer ear (auricle); prevent accidental injuries from flying objects in industrial activities, from contact during physical activities, or from excessive heat or fire; prevent moisture or solid particles from entering the ear canal; reduce noise to acceptable levels so signals are heard better through earphones; and protect the outer ear from direct sun rays. The purpose for wearing ear protectors determines the characteristics of the device to be used.

Hearing protectors. The most common use of ear protectors is to protect hearing from excessive noise; therefore "hearing protector" is a most appropriate term. Dry cotton has been used for many years as a form of hearing protection, but unfortunately this material provides very limited protection against the intrusion of excessive noises. Developments in plastics have greatly improved the amounts of protection that may be achieved. Comfort is one of the major factors in the successful use of all devices. The amount by which a device reduces or attenuates the magnitudes of noises reaching the structures of the inner ear is also an important consideration. Advances in materials used to construct hearing protection devices have led to significant improvements in both attenuation and comfort for users. The use of hearing protectors can prevent both temporary (auditory fatigue) and permanent noise-induced hearing losses, and may actually enhance a person's ability to hear in the presence of excessive noise. *See* HEARING (HUMAN).

Hearing protectors that insert into the opening of the ear canal are usually referred to as earplugs. Those that enclose the structures of the outer ear are usually called earmuffs. The manner in which each is constructed and used may differ considerably. A variety is found among earplugs: premolded plugs designed to accommodate the anatomic characteristics of most human ears; custom-molded devices that mirror the anatomic contours and features of an individual ear; expandable material which is compressed into a cylinder and inserted into the opening of the canal, where it expands to make a comfortable seal; fibrous material partially formed by the user and inserted into the opening of the ear canal; and earcaps that are held in place against the opening of the ear canal by pressure from a light-weight headband. Earmuffs usually consist of a molded cup completely encasing the outer ear and fitted with a compliant cushion that rests on the head in the area surrounding the ear. Earmuffs are held firmly in place by pressure from a headband which may be placed over the top of the head, under the chin, or behind the neck. Earmuffs may also be attached to a hardhat or helmet.

Attenuation measurement. Various schemes have been devised to measure and express the noise reduction or attenuation provided by different types of hearing protectors. The most common method used for expressing amounts of attenuation compares the sound reduction that occurs when the device is worn with values obtained when the ears are not protected. Studies that have compared attenuation under laboratory conditions with that achieved in actual usage consistently reveal that users do not typically obtain the amounts of protection measured in controlled laboratory situations. The most common method for expressing values of attenuation is the noise reduction rating (NRR). Generally, NRR values range from 0 to approximately 30, with higher numbers indicating greater amounts of noise reduction (see **illus.**). *See* LOUDNESS.



Noise reduction ratings (NRR) for various types of personal hearing-protection devices. A total of 242 devices were evaluated. (*After B. L. Lempert, Compendium of hearing protection devices, Sound Vibrat., 18(5):26–39, 1984*)

Protection from water. People frequently wear earplugs to prevent water from entering the ears while taking a shower or swimming. While few earplugs totally exclude water from entering the ear canals, especially if the head is completely immersed, some, if properly fitted, can significantly impede the passage of water. The pressure exerted by the water on the outer surface of the plugs increases dramatically the deeper one goes below the surface, making it even more difficult to exclude the water totally.

Noise identification. Millions of people possess some degree of noise-induced loss which could have

been prevented by wearing hearing protection during encounters with excessive noise. Controlling such excessive exposures requires that hazardous noises be identified. A fairly simple approach is to consider the following rule: if the noise encountered would cause one to use a loud voice to communicate at a distance of approximately 1 ft (0.3 m), the noise should be avoided or the levels reaching the inner ears should be reduced by the use of hearing protection. *See* ACOUSTIC NOISE. Donald C. Gasaway

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Early modern humans

The earliest representatives of people anatomically similar to living humans evolved from more archaic humans approximately 100,000 years ago. The process by which they emerged from and eventually replaced those late archaic humans remains unclear. It is likely that they evolved locally from such predecessors in northeastern Africa. Over the succeeding 50,000 years, their range expanded and contracted with changing global climatic cycles to include at times the Mediterranean Near East and portions of northern Africa. Early modern humans and their biology and way of life, therefore, initially had little advantage over late archaic humans. *See* NEANDERTALS.

Subsequently, starting around 50,000 years ago, the descendants of these earliest modern humans were able to expand their geographical range. They spread into regions occupied by geographical groups of late archaic humans, in some cases admixing with those local populations and in other regions displacing them. The spread of early modern humans appears to have taken place initially across northern Africa to Morocco and southward through eastern Africa to southern Africa. At the same time, prior to 40,000 years ago, they appear to have spread eastward across southern Asia, through Indonesia, and across open water into greater Australia. Between 40,000 and 30,000 years ago, these early modern humans dispersed northward across Eurasia from the Mediterranean to the Pacific, and westward across Europe, reaching the Atlantic peripheries of Europe as late as 30,000 years ago. The last of these groups, the early modern humans of western Europe, were previously referred to as the Cro-Magnon people, after the site in France where their remains were first discovered in 1868.

This scenario for the emergence and spread of modern humans is based on evidence from the human fossil record (documenting the earliest modern humans in different regions), living human genetic diversity (similarities among humans, indicating a generally recent shared ancestry), and recent human anatomical diversity (requiring considerable time to become established). Taken together, these lines of evidence highlight the complex and geographically variable nature of this evolutionary period, the emergence of modern humans.

Early modern humans were biologically the same as modern peoples and would blend in with living peoples. They differed from living people primarily in their tendency to have a rugged, athletic build. This is evident in the prominent muscle attachments and the structural reinforcements of the limb bones. Given their mobile hunting and gathering existence, their legs and feet were especially strongly built. In addition, they had slightly larger teeth than those of recent humans, providing greater resistance to the abrasion of eating and cooking without utensils.

These populations were generally tall, males being 175–180 cm (5 ft 10 in.) and females 160–165 cm (5 ft 4 in.) on average. From about 20,000 years ago until the twentieth century, few human populations achieved such large statures. As a result of these tall bodies and muscularity, their brains were relatively large, averaging about 1500 cm^3 (90 $in.^3$) as opposed to averages of about 1350 cm^3 (80 $in.^3$) common for recent humans. Yet, when their brain sizes are scaled against estimated body weights, their brains were relatively the same size as those of living humans.

Early modern humans were successful hunters and gatherers, occupying most of the inhabitable regions of the Old World. Although frequently portrayed as big game hunters, they lived by hunting small to medium-size animals, especially antelopes, deer, goats, and occasional horses and cattle, and by gathering wild plants, fish, and shellfish for food. They were sufficiently successful that the levels of biological stress from famine or injury, as indicated by the lesions on their skeletons and teeth, are among the lowest known for any prehistoric human group.

Their effectiveness as hunters and gatherers was due in part to their technology. They developed elaborate stone tool technologies, producing long blades that became blanks for tools with replaceable cutting edges and points. This advance was made possible by the use of bone, antler, and wood for carefully made hafts and handles for the sharp stone edges. They were also the first to fire clay into ceramics, and they wove carrying bags with a variety of techniques. Yet, their ability to live effectively as hunters and gatherers depended upon their extensive knowledge of the environment, so that they could harvest wild game and plants rather than seek them opportunistically. This knowledge was communicated through the first elaborate symbolic systems known, which consisted of a variety of geometric notational systems and the first representational art. They were also the first humans to commonly wear jewelry, and hence to modify their personal social images, suggesting more complex social roles than were previously known.

Although these behavioral advances are associated with early modern humans, most of them appear

only after about 50,000 years ago and hence are associated with the dispersal of modern humans. What behavioral advances, if any, were associated with the earliest modern humans remains unclear. However, their fossil remains suggest that even they were less stressed and more efficient as hunter-gatherers than their archaic predecessors and contemporaries. *See* FOSSIL HUMANS.

Erik Trinkaus

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Earphones

A class of energy transducers capable of receiving alternating current and generating acoustic waves resembling very closely the characteristics of that current. The movement of an element (diaphragm) is accomplished by magnetic attraction, electrostatic attraction, or the piezoelectric effect (the expansion or contraction of certain crystalline substances in response to electric charges). Earphone systems include the driver element with its diaphragm and arrangements for magnetic flux, or electrostatic or direct electric charge, plus a casing, one or more acoustic cavities and ports, acoustic damping and insulation, and some arrangement for coupling the driver to the human ear. The wiring connecting the precedent amplifier to the driver, which may be incorporated into the earphone system, may in modern systems be a complicated circuit which feeds, for each of two stereophonic channels, some part of the current to each ear. The time delays and energy ratios at the two earphones (each of which may contain two drivers) can be appropriately adjusted so that the listener is given the illusion that the sound sources are not "within the head" but are externalized appropriately in all three planes of space. The various types of earphones are described below. *See* TRANSDUCER.

Magnetic. In this type, a permanently magnetized diaphragm is moved in and out by an electromagnet energized by alternating current. Early models had two poles attracting and repelling a relatively heavy metal diaphragm with complex vibrational modes, the frequency response being limited to about 300–2500 Hz. In one development, the heavy metal diaphragm was replaced by a thin sheet of aluminum held by a ring of magnetic metal. The coil through which the signal current traveled had to move only the relatively low-weight ring, a greatly increased efficiency and with improved frequency response.

Dynamic. These earphones are actually small dynamic loudspeakers. In some, a small coil fed by the sound source is bonded to the membranous diaphragm (**Fig. 1**). Alternating current in the coil thus drives the diaphragm toward and away from a permanent magnet in the rear of the casing. In another configuration, the coil is relatively large and is attached to the diaphragm only at its edge; the diaphragm is a molded dome with a rolled edge and moves freely in the magnetic field. Such earphones are standard for high-fidelity communications.

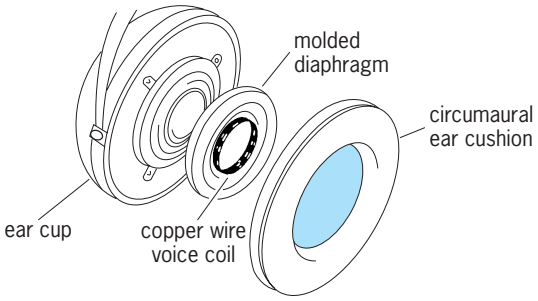

Fig. 1. Exploded view of conventional dynamic earphone. (*After P. Milton, Headphones: History and Measurement, Audio, 26(5):90–94, May 1978*)

Miniature magnetic. A common earphone is a small unit of which the output port fits snugly into a plastic olive in the ear canal (**Fig. 2**). These are widely used with small radios, with the more powerful hearing aids, by television commentators, in business transcription devices, and in many other communications situations. A coil fed by the signal current is wrapped around a pole and imparts in-and-out motion to the diaphragm proportional to the alternating current flow.

Another configuration (**Fig. 3**) is incorporated inside the case of most ear-level and all-in-the-ear hearing aids. A metal armature has a free end between the two poles of a magnet. When alternating current is flowing in one direction, the armature at the free end is turned into a south pole and moves toward the north pole of the permanent magnet. As the current changes direction, the free end is turned into a north pole and the armature moves in the other direction. The armature moves a diaphragm, creating sound which is ported into the ear canal. *See* HEARING AID.

The diaphragm can be of any light material, not necessarily magnetic, and can be formed into an efficient shape for producing and directing an acoustic wave. Such miniature earphones are now made so efficient as to compete favorably with much larger types with respect to acceptably flat frequency

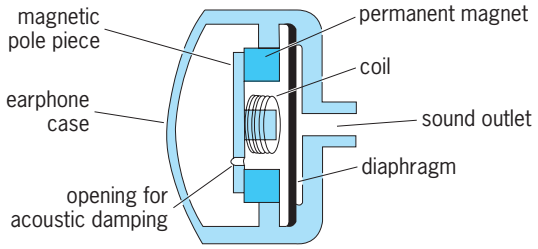

Fig. 2. Cross section of external-type hearing aid earphone. (*After W. R. Hodgson and P. H. Skinner, eds., Hearing Aid Assessment and Use in Audiologic Habilitation, Williams and Wilkins, 1977*)

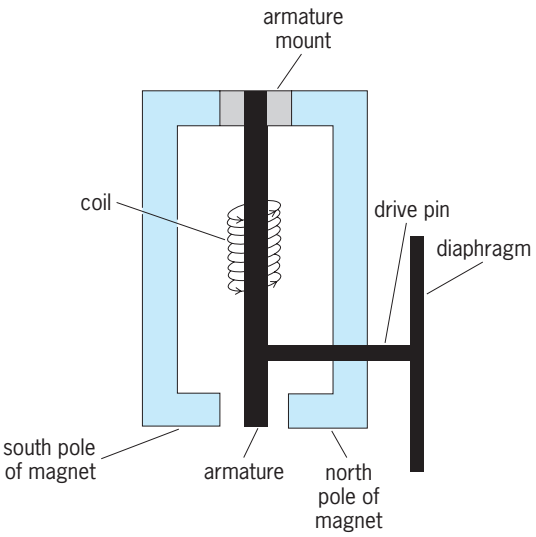

Fig. 3. Cross section of internal-type hearing aid earphone. (*After W. R. Hodgson and P. H. Skinner, eds., Hearing Aid Assessment and Use in Audiologic Assessment, Williams and Wilkins, 1977*)

response in the octave above 10 kHz and with respect to sound power levels in the ear canal with acceptably low total harmonic distortion.

Electrostatic. Efficiency is increased if the mass of the diaphragm is reduced to a minimum. A thin (2.5–12 micrometers) metallized plastic film can be used, on which a large constant electrostatic charge is placed by an auxiliary unit, and the motion of the diaphragm is controlled by the audio signal impressed on perforated wire mesh plates on either side (the push-pull arrangement reduces second-harmonic distortion). The assembly is mounted in a relatively large cavity and coupled to the head by a circumaural cushion (that is, one which completely surrounds the auricle). Such earphones (**Fig. 4**) are light and comfortable and have excellent frequency-response and transient-distortion characteristics.

Dynamic-electrostatic. The diaphragm of a dynamic-electrostatic (or orthodynamic) earphone (**Fig. 5**) is a permanently polarized electret of fluorocarbon. Consequently the need for an added source of polarization voltage (a drawback inherent to electrostatic earphones) is eliminated. In one development, certain problems of such a system are avoided by separating the functions of the diaphragm and the electret. The diaphragm can be of the thinnest (2.5 micrometers) metallized polyester, stretched between two electrets in push-pull. The diaphragm is matched in acoustic impedance very closely to the surrounding air, and the audio signal can be led directly to the electrets. *See* ELECTRET TRANSDUCER.

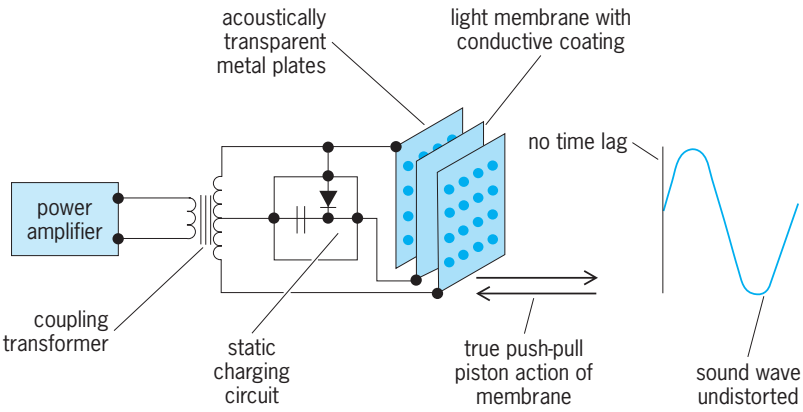

Fig. 4. Block diagram of electrostatic earphone. (*After H. Souther, An Adventure in Headphone Design: The Model ESP-6 Electrostatic Headphone, Koss Electronics Inc., Cat. 111, 1969*)

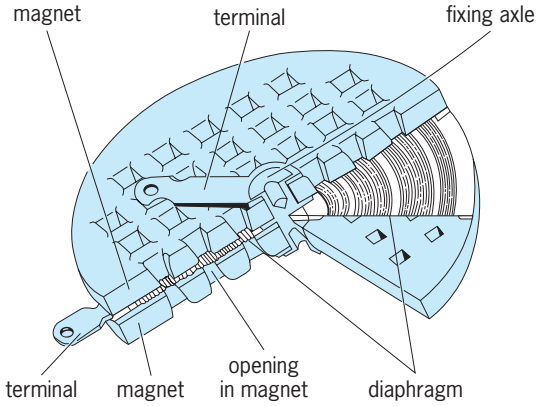

Fig. 5. Construction of orthodynamic earphone driver. (*After P. Milton, Headphones: History and Measurement, Audio, 26*(5):*90–94, May 1978*)

Piezoelectric. Certain crystalline substances expand and contract when alternating voltage is applied. In some piezoelectric earphones, a crystal element is coupled mechanically to the center of a small (about 2.5 cm diameter) cone. Such earphones can be lightweight and cheap and may be acceptable for speech communication.

Substances other than quartz crystals, notably high-polymer films such as vinylidene fluoride, will expand and contract when an alternating current is applied. A thin sheet of such material can be stretched by a factor of 4, and aluminum vapor deposited over it. If a high dc voltage is now applied, and removed, the sheet will become and remain piezoelectric. Two sheets are bonded together back to back, and when alternating current is applied between the metallized faces the sheet will expand and contract. If a rectangular sheet is slightly curved and mounted rigidly top and bottom, much as a playing card might be grasped between the fingers, the movement in response to alternating current will be in and out in proportion and an acoustic wave will be created (**Fig. 6**). Earphones on this principle have desirable characteristics of stability, acoustic response, and cost, but require more amplification than an electret or other earphone to operate as efficiently. *See* PIEZOELECTRICITY.

Real-ear response. An earphone system contains components which are in some respects as important as the driver. The configuration of the acoustic ports, for example, or the acoustic or other damping built into the system will help determine the

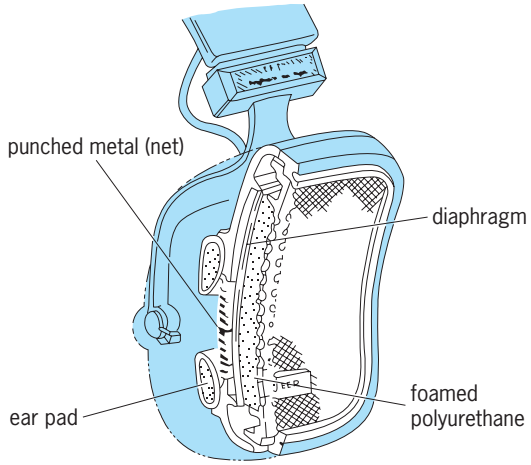

Fig. 6. Cutaway view of rear of high-polymer piezoelectric earphone. (*After P. Milton, Headphones: History and Measurement, Audio, 26(5):90–94, May 1978*)

eveness of the frequency response, while the cushions, whether circumaural or supraaural (fitting against the auricle), help determine the low-frequency real-ear response.

The effects of changes in these elements as reflected in acoustic differences at the eardrums of actual persons are not easy to determine. It is not difficult to couple an earphone system to a microphone by means of a standard cavity, perhaps incorporating a flat plate for the larger circumaural systems, but such a practice is useful primarily for the manufacturer's quality control, and has limited relevance to the performance of the system on even the average human head. The reference equivalent threshold sound-pressure level of a certain earphone system, expressed in voltage applied as frequency changes, has to be determined by having a panel of normal-hearing listeners loudness-balance that system at each frequency against another (arbitrarily chosen) reference system.

The acoustic impedance of the several elements of the external ear, head, and torso have been measured, and artificial heads incorporating the results of such measurements are available. These devices allow fairly successful predictions as to how an earphone system will perform on an anthropometrically average person. *See* ACOUSTIC IMPEDANCE.

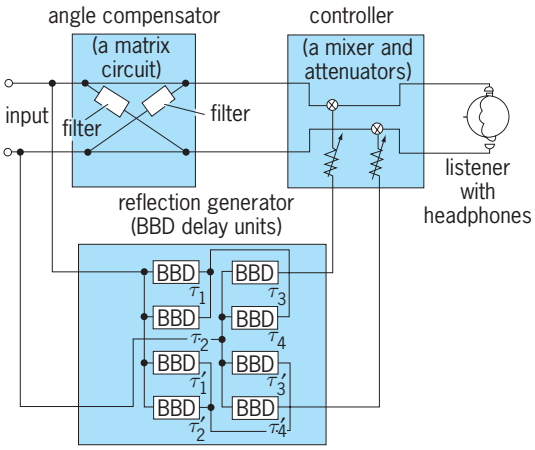

Fig. 7. Block diagram of projected sound localization earphone system for stereophonic sources. BBD-bucket brigade device. (*After N. Sakamoto, T. Gotoh, and Y. Kimura, On "out-of-head" localization in headphone listening, J. Audio Eng. Soc., 24:710–716, 1976*)

Realistic simulation. Any single-channel recording of a real acoustic event will, when played back to an earphone in one ear, sound "in the head." Even stereo recordings from two channels played back to earphones on two ears, while furnishing the illusion of movement, still are not externally localized by the listener. Externalization is improved if some of the signal from the left channel is time-delayed and applied also to the right earphone (the same, of course, for the right channel to the left earphone), and it is also improved if a realistic ratio is achieved between the acoustic energy density ratio from the "direct" versus the "indirect" sound sources (as from reflective walls). Stereophonic earphone systems have been built which incorporate these time delays and ratios, and which furthermore feed the signal from the right channel of an artificial-head stereophonic recording to the left earphone (and the left channel to the right earphone) through frequency filters which simulate the differential acoustic effects at various frequencies of the head and external ears in the original recording situation. Thus the eardrums under earphones are presented with the exact acoustic conditions generated by a loudspeaker or other sound source in an actual room, and such earphone systems very materially advance the important psychoacoustic feature of externalization of sound and of acoustic realism generally (**Fig. 7**). *See* LOUDSPEAKER; MICROPHONE; SOUND-REPRODUCING SYSTEMS; STEREOPHONIC SOUND.

J. Donald Harris

Earth

The third planet from the Sun and the largest of the four inner, or terrestrial, planets. The Sun is an average-sized, middle-aged star situated toward the outer edge of one of the spiral arms of the Milky Way Galaxy. So far as is known, Earth is unique in the solar system in having life. Whether life exists in the universe beyond the solar system is unknown. However, recent discoveries have proven that there are planets in orbit around many nearby stars, so there are possibly billions of Earth-like planets in the universe where life may yet be discovered. *See* GALAXY, EXTERNAL; MILKY WAY GALAXY; PLANET; SOLAR SYSTEM; STAR; SUN; UNIVERSE.

Earth has one natural satellite, the Moon. Otherwise, Earth's nearest neighbors in space are Venus, which is about 108×10^6 km (67×10^6 mi) from the Sun, and Mars, about 228×10^6 km (141×10^6 mi) from the Sun. Earth is about 150×10^6 km (93×10^6 mi) from the Sun. Distances between planets are so immense that it is difficult to comprehend them. To put the solar system into perspective, think of the Sun as a basketball. The planet closest to the Sun, Mercury, would be the size of a grain of sand slightly

less than 1 mm in diameter and 12 m from the basketball. Venus, the second planet out, would be a granule 2.4 mm in diameter, 22 m away, and Earth would be a granule 2.5 mm in diameter, 30 m from the basketball. *See* MARS; MERCURY (PLANET); MOON; VENUS.

Motions. A person standing on the surface of Earth undergoes a complex motion through space. The motion has two principal components, one arising from the orbital motion of Earth around the Sun, the other from the rotation of Earth around its axis.

Orbital motion. Earth completes an orbit around the Sun in 365 days, 5 h, 48 min, 46 s; the orbit defines the length of the year. The length of the day is determined by the period of Earth's rotation about its axis. The fact that the year is not a whole number of days has affected the development of the calendar. *See* CALENDAR.

The orbit is close to circular but is actually elliptical with the Sun at one focus. Earth's closest approach to the Sun (perihelion), which happens about January 3, brings Earth to within 147×10^6 km (91.5×10^6 mi); Earth is at aphelion (farthest from the Sun) about July 4, when Earth is 152×10^6 km (94.5×10^6 mi) away. The average velocity of Earth in its orbit is 107,000 km/h (66,600 mi/h), or 29.6 km/s (18.5 mi/s). Because the orbit is elliptical, Earth's velocity varies systematically, being greatest at perihelion and least at aphelion. The changing velocity affects measurements of solar time. *See* CELESTIAL MECHANICS; ORBITAL MOTION.

Rotation. Earth rotates on its axis once each day. The axis of rotation is perpendicular to the Equator, and the Equator is inclined at about 23.5° to the plane of Earth's orbit around the Sun. As Earth moves in its orbit, the north spin axis, or north geographic pole, points in the direction of the star Polaris, making it the North Star or polestar. One result of the tilt of the Equator relative to the orbital plane is that different parts of Earth receive differing amounts of sunlight through the year; this is the primary cause of seasons. *See* EQUATOR; POLARIS; ROTATIONAL MOTION.

Earth's spin axis has not always pointed in the direction of Polaris. Earth is not a sphere; it bulges at the Equator so that the equatorial radius is 21 km (13 mi) greater than the polar radius. The gravitational attractions of the Sun and Moon on the equatorial bulge cause the spin axis to slowly trace out a cone in space—that is, as time passes, the north spin axis will no longer point toward Polaris. This slow conical motion of the spin axis is called precession, and a complete cycle requires 25,800 years. Superimposed on the precession is a much smaller fluctuation, called nutation, that results from the movement of the Sun twice a year, and the Moon twice a month, across the plane of the celestial equator. *See* EARTH, GRAVITY FIELD OF; PRECESSION OF EQUINOXES.

Earth's period of rotation, the day, is used to define the second, which is the basic unit of time. The length of the day varies irregularly owing to tidal friction, motions in Earth's core, and seasonal atmospheric circulation. For this reason, the second was defined as 1/86,400 of a mean solar day averaged over a year. To avoid the obvious problems in determining the length of the mean solar day, the second is now defined atomically. *See* DAY; TIME.

Earth's speed of rotation is a maximum in late July and early August, and a minimum in April; the difference in the length of the day is about 0.0012 s. Since about 1900, Earth's rotation has been slowing at a rate of about 1.7 s per year. In the geologic past, Earth's rotational period was much faster. Daily, monthly, tidal, and annual growth rings on fossil marine organisms reveal that about 350×10^6 years ago (Middle Devonian Period) the year had 400–410 days, and 280×10^6 years ago (Pennsylvanian Period) the year had 390 days. These observations are very close to the calculated values of the effect of tidal friction. The slowing of Earth's rotation is caused by tidal friction between the sea floor and the ocean water. The Moon is the main cause of tides, and the total rotation energy in the Earth-Moon system is conserved, so that the energy lost by Earth is gained by the Moon. This causes the Moon to move farther from Earth, and this in turn lengthens the period of the Moon's revolution. *See* DEVONIAN; EARTH ROTATION AND ORBITAL MOTION; PENNSYLVANIAN; TIDE.

Satellites. The Moon is Earth's only natural satellite. The Moon's mass is 1/81.3 of Earth's, and its average distance from Earth is 383,403 km (238,247 mi). The center of mass of the Earth-Moon system is within Earth, about 4645 km (2886 mi) from Earth's center, and it is about this point that the Earth-Moon system revolves.

The Moon's period of revolution and rotation is 27 days, 7 h, 43 min, 11.5 s. The orbital plane of the Earth-Moon system is inclined to the orbital plan of the Earth-Sun system at an angle of 5°8′33″. At times the Sun, Earth, and Moon are in a line, and an eclipse of the Sun or the Moon occurs. *See* ECLIPSE; SATELLITE (ASTRONOMY).

Size, shape, mass, and density. Earth is an oblate spheroid. The mean equatorial radius is 6378.139 km (3963.37 mi), and the polar radius is 6356.779 km (3950.10 mi), the difference being 21.360 km (13.27 mi).

Earth's mass is 5.976×10^{27} g (0.2108×10^{27} oz), being the sum of 5.974×10^{27} g (0.2107×10^{27} oz) for solid Earth, 1.4×10^{24} g (0.049×10^{24} oz) for the ocean, and 5.1×10^{21} g (0.18×10^{21} oz) for the atmosphere. Earth's average density is 5.518 g/cm^3, which is just about double the density of the common rocks that form at Earth's surface, indicating that Earth's interior is more dense than the surface. Seismic studies have confirmed that Earth is layered both compositionally and mechanically (**Fig. 1**). *See* ATMOSPHERE; DENSITY; EARTH INTERIOR; OCEANOGRAPHY; SEISMOLOGY.

Internal structure. The deepest compositional layer is the core, which is divided into a solid inner core and a liquid outer core. Both the inner and outer core have the same composition, believed to be nickel-iron plus a small amount of lighter elements such as sulfur and silicon. Electric currents moving in the molten metal outer core are believed to be the origin

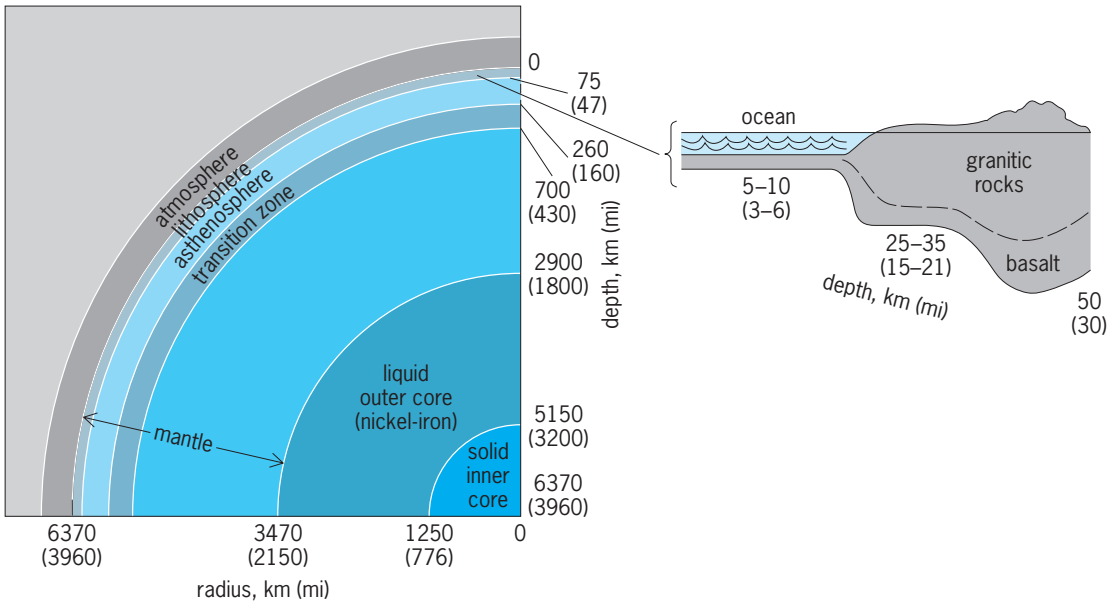

Fig. 1. Principal layers of Earth.

of Earth's magnetic field. Above the core is the mantle which, on the basis of density of rare rock samples brought up from deep in the mantle in kimberlite pipes, and other evidence, is believed to be composed of silicate minerals, and in particular olivine and pyroxene. A rock composed largely of olivine and pyroxene is called a peridotite. *See* OLIVINE; PERIDOTITE; PYROXENE; SILICATE MINERALS.

Within the mantle there are two prominent seismic discontinuities, one at a depth of about 400 km (248 mi), the other at a depth of about 670 km (416 mi). The seismic discontinuities are marked by pronounced changes in the velocities of seismic waves. The 400-km discontinuity is presumed to be due to the breakdown of pyroxene to garnet plus stishovite, and the transformation of olivine to a spinel structure with a consequent increase in density of about 5%. The 670-km seismic discontinuity is thought to be due to the breakdown of olivine spinel to perovskite and periclase, and the transformation of garnet to a perovskite structure. The lower mantle, from 670 km to the core-mantle boundary at a depth of 2885 km (1793 mi), does not exhibit any pronounced seismic discontinuities. *See* GARNET; PEROVSKITE; SPINEL; STISHOVITE.

Above the mantle is Earth's crust, and between the crust and the mantle there is a pronounced seismic discontinuity known as the Mohorovičić discontinuity, or Moho. The crust is of two kinds, both of which are less dense and compositionally different from the peridotitic mantle below. Beneath the ocean the crust is basaltic in composition and about 8 km (5 mi) thick. The crust beneath the continents is granitic in composition and averages 35 km (21.7 mi) in thickness but ranges up to 80 km (49.7 mi), as beneath Tibet. The oceanic crust is geologically young because it is continually created and destroyed through the process of plate tectonics. No part of the oceanic crust that is older than about 180×10^6 years has yet been discovered. The continental crust is much older than the oceanic crust. Continental rocks as old as 4×10^9 years have been discovered in Canada, and the fact that they are highly deformed indicates a long and eventful history. *See* EARTH CRUST; GRANITE; PLATE TECTONICS.

Mechanical layering. The internal temperature of Earth increases from an average of 15°C (59°F) at the surface to about 5000°C (9032°F) at the center of the core, approximately the same temperature as the surface of the Sun. Measurements show that the thermal gradient beneath the ocean is about 13°C/km (69.7°F/mi), while that in the continental crust is about 6.7°C/km (51.4°F/mi). The combined influence of temperature and pressure are such that rocks in the crust and upper mantle are mechanically strong and rigid up to a temperature of about 1300–1350°C (2372–2462°F), equivalent to a depth of about 100 km (62 mi) beneath the oceans and 200 km (124 mi) beneath the continents. The layer of mechanically strong rocks, which includes the uppermost 92 km (57 mi) of the mantle beneath the oceanic crust and the uppermost 165 km (103 mi) of the mantle beneath the continents, is called the lithosphere. Below the lithosphere, rocks lose strength and can be readily deformed because temperatures and pressure almost reach melting conditions of mantle rocks. A layer of weak rocks about 100 km thick and known as the asthenosphere underlies the lithosphere. The mechanical strength of the lithosphere is such that heat moves through it by conduction, but the strength of the asthenosphere is such that heat can move through it by convection. Beneath the asthenosphere both temperature and pressure increases with depth, but the temperature gradient is only about 0.5°C/km (1.4°F/mi) because it is determined by convection. As a result, the strength of rocks in the middle and lower mantle slowly increases, but at all depths below the

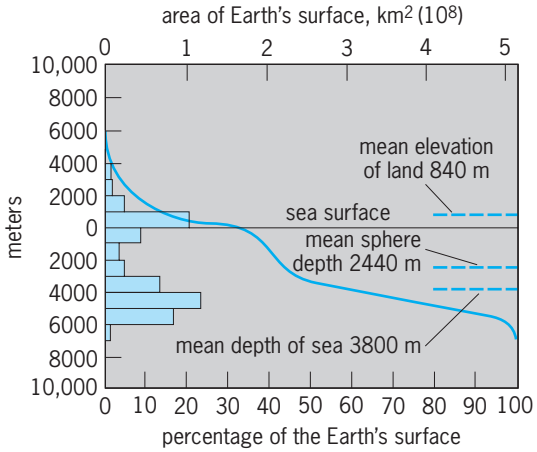

Fig. 2. Hypsographic curve of land surface elevation. 1 m = 3.3 ft; 1 km² = 0.4 mi². (*After E. Kossinna, Die Tiefen des Weltmeeres, Veröff. Inst. Meereskunde Univ. Berlin, Neue Folge A. Geogr.-naturwiss., 9:1–70, 1921*)

lithosphere mantle rocks are capable of flow so that heat moves outward from the core to the base of the lithosphere by convection. *See* ASTHENOSPHERE; CONDUCTION (HEAT); CONTINENTS, EVOLUTION OF; EARTH, CONVECTION IN; LITHOSPHERE.

Crustal structures and surface features. The surface of solid Earth has a bimodal distribution of elevations (**Fig. 2**). If the water from the ocean could be removed, it would be apparent that continents stand high (average elevation is 840 m or 2755 ft above sea level), while the ocean floor sits low (average elevation is 3800 m or 12,464 ft below sea level). This difference in elevation arises because rigid lithosphere floats on the weak asthenosphere, and because the density of oceanic lithosphere (that is, lithosphere capped by oceanic crust) is greater than the density of continental lithosphere.

On the continents, mountain belts are the most dramatic features. They range in elevation from Mount Everest, 8848 m (29,030 ft), in the Himalaya Mountains to older, deeply eroded ranges that are now barely above sea level. Granitic and metamorphic rocks are generally exposed in the cores of mountain ranges. The overlying rocks that cover most of the Earth's surface are sedimentary, mainly of shallow marine origin, that may or may not have been deformed. The deformation is the result of compression and tension that causes folding and faulting, and may be accompanied by intrusion and metamorphism. Movements and collisions of tectonic plates are the principal cause of mountain building. Mountains generally are formed over several tens of millions of years. The rocks deformed in the process are generally marine sedimentary rocks formed along the margins of continents. *See* DEEP-MARINE SEDIMENTS; MARINE SEDIMENTS; METAMORPHIC ROCKS; OROGENY; SEDIMENTARY ROCKS.

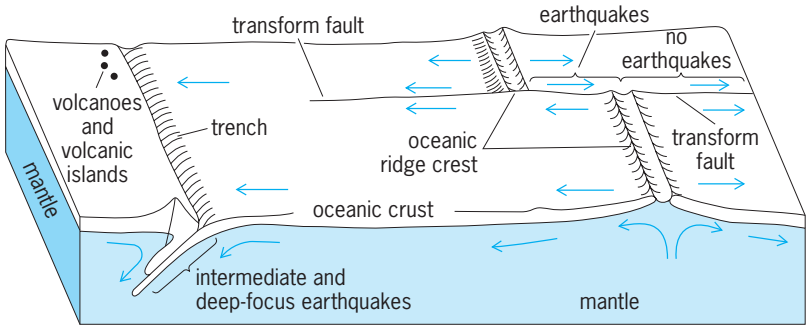

Fig. 3. Sea-floor spreading. (*After G. Gross, Oceanography, 3d ed., Charles E. Merrill, 1976*)

Some topographic mountains are formed by volcanoes. The most striking are long chains of andesitic stratovolcanoes, such as the Aleutian and Andean chains, formed as a result of melting in the mantle induced by sinking tectonic plates. *See* VOLCANO.

Much of the surface of the continents is covered by a thin veneer of sedimentary rocks. Where the underlying rocks of the plains and hills that make up most of the continents are exposed, the rocks and their structures are similar to those found in most mountain ranges. This leads to the theory that continents are formed by the deeply eroded remnants of earlier mountain ranges, and that today's continents have been assembled by accretion and by plate collisions.

The topographic features underlying the oceans are similarly diverse and reveal more evidence of a dynamic Earth. The continental shelf, an area covered by shallow water, generally less than 150 m (500 ft) deep, surrounds the continents at most places. Such areas are generally underlain by continental, granitic rocks, and are submerged parts of the continents. Continental slopes are the transition between the continental shelf and the ocean floors. Their tops are generally less than 150 m below sea level, and they slope down to about 4400 m (14,000 ft). They are narrow, steep features, with slopes generally between 2 and 6°, but some are up to 45°. They are generally underlain by thick accumulations of sedimentary rocks. *See* SEA-FLOOR IMAGING.

Submarine trenches and their associated volcanic island arcs are formed as a result of a tectonic plate of lithosphere sinking into the mantle beneath the edge of an overriding plate. The deepest place on Earth is in the Mariana Trench, 11,022 m (36,152 ft) below sea level.

The ocean floor is the most widespread surface feature of Earth. Beneath an average of 4.4 km (2.75 mi) of seawater are about 2.3 km (1.4 mi) of sedimentary rocks with some intercalated basalt, and below that is the oceanic crust, consisting of 4–6 km (2.5–3.7 mi) of basaltic rocks. Interrupting the ocean floor at many places are submarine mountains formed by basaltic volcanoes. Some of these volcanoes are very large and form oceanic islands such as the Hawaiian Islands. At other places, vertical movements of the ocean floor are revealed by sunken islands whose flat tops were formed by wave erosion (guyots) and coral atolls.

The ocean floor rises gradually to the mid-ocean ridges, a more or less continuous feature through all the oceans with some branches and offsets. The

ridges range between 480 and 4800 km (300 to 3000 mi) wide and are much more rugged and irregular than the ocean floors. They rise about 3000 m (10,000 ft) above their base on average. Lines of parallel volcanoes, steep scarps, and a central valley mark the mid-ocean ridges. The central parts of the generally more or less symmetrical ridges are the most active volcanically and seismically. *See* MARINE GEOLOGY; MID-OCEANIC RIDGE.

The source of the heat energy that deforms the surface rocks is the crystallization of the liquid core and the decay of radioisotopes within the whole Earth. The energy released at the bottom of the mantle may be the cause of convection in the mantle, which in turn causes plate tectonics and so deforms the crust, but the actual processes are probably much more complex and are not yet clearly understood. *See* RADIOISOTOPE (GEOCHEMISTRY).

Plate tectonics. New oceanic lithosphere capped by basaltic crust is created at the mid-ocean ridges, and this newly formed plate moves away from the ridges (**Fig. 3**). The tectonic plates formed in this way may carry continents on them, and are the mechanism of continental drift. Paleomagnetic data from the continents indicate that the continents have moved relative to each other. The tectonic plates capped by basaltic crust plates are consumed at the trench–volcanic island arc areas. *See* PALEOMAGNETISM; SUBDUCTION ZONES.

As well as the ridge and the trench, a third type of plate boundary occurs where two plates slide past each other at a transform fault. Moving plates may collide in several ways (**Fig. 4**). Such collisions account for the deformed rocks found in the crust. *See* TRANSFORM FAULT.

The evidence for continental drift in the geological past includes matching of rock types, ages, fossils, climates, and structures (mountain ranges), as well as the paleomagnetic data. Evidence showing or suggesting present movements consists of shallow earthquakes along mid-ocean ridges and transform faults that offset them; deep earthquakes associated with deep-sea trench–volcanic island arc areas; direct measurement of movement; volcanic activity at mid-ocean ridges; and volcanic activity at trench–island arc areas. *See* GEODESY.

Atmosphere. Earth's temperature and gravitation are such that an atmosphere is present. The major constituents are nitrogen and oxygen. The atmosphere, especially oxygen, and the presence of water, both at the surface and in the atmosphere, make life possible. Precipitation, mainly rain, results in running water such as streams and rivers on the continents. Running water is the main cause of erosion of the continents, and most of the landscapes have been eroded by water, although some are eroded by wind or ice (glaciers). *See* EROSION; GLACIOLOGY; GRAVITY; NITROGEN; OXYGEN; WATER.

The atmosphere shields Earth from most meteorites. Meteors are caused by frictional heat when such high-velocity objects move through the atmosphere. A few of these objects pass through the atmosphere and strike Earth, causing craters. Other bodies, such as the Moon, are not protected by an atmosphere and so are struck more frequently by meteorites. The craters formed on the Moon are preserved, because without an atmosphere almost no erosion occurs. *See* METEORITE.

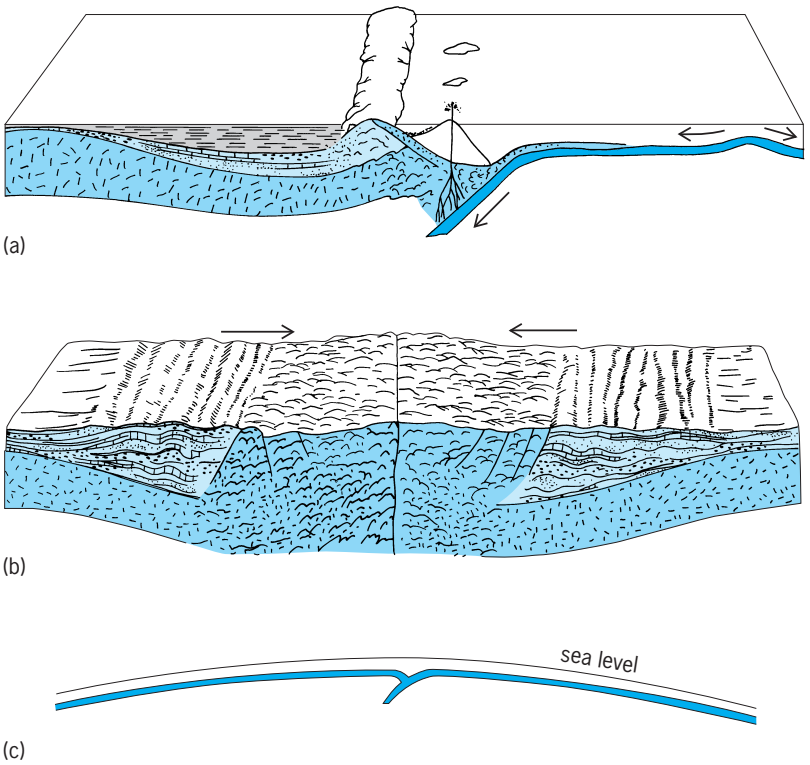

Fig. 4. Plate collisions: (*a*) continent-ocean plate collision; (*b*) continent-continent plate collision; (*c*) ocean-ocean plate collision. (*After R. J. Foster, Physical Geology, 3d ed., Charles M. Merrill, 1979*)

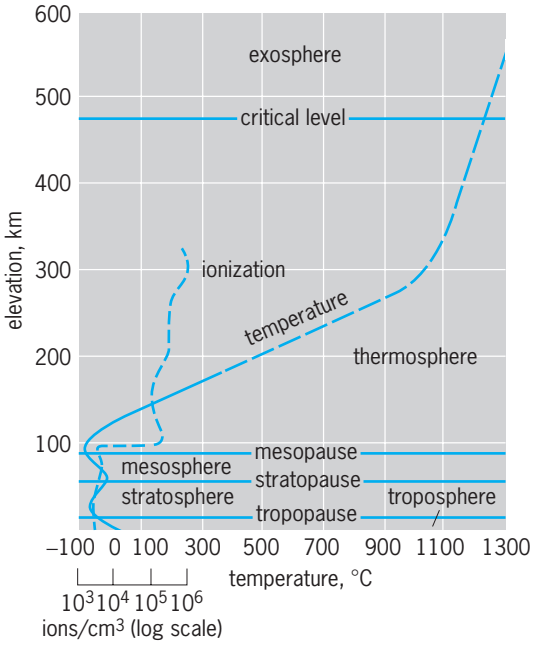

Fig. 5. Structure of the atmosphere. The log scale applies only to the ionization curve. 1 km = 0.6 mi; °F = (°C × 1.8) + 32.

A thin ozone layer in the atmosphere also shields the Earth from lethal ultraviolet radiation from the Sun. The structure and composition of the atmosphere is shown in **Fig. 5**. *See* OZONE.

Age. Earth, along with the rest of the solar system, is believed to have formed about 4.55×10^9 years ago. This age is determined by dating radioactive isotopes in meteorities. Meteorites are believed to be fragments produced by collisions among small bodies formed by the same process that created the solar system. Theoretical studies of the Sun and other studies of radioactive isotopes also suggest a similar age. *See* EARTH, AGE OF.

Brian J. Skinner

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Earth, age of

An estimate of the age of the planet Earth based on the lead isotope systematics. The first meaningful estimate was calculated from the average isotopic composition of terrestrial lead (Pb) available at the surface of the planet. The feasibility of the calculation stems principally from the unique circumstance that two isotopes of the element uranium (^{235}U and ^{238}U) decay to two daughters that are also isotopes of another element, lead (^{207}Pb and ^{206}Pb, respectively). This eliminates the requirement of determining the planetary budget of either of the two elements. The total number of ^{206}Pb and ^{207}Pb atoms are related to their respective parents by the equation of radioactive decay, expressed as Eq. (1) for lead-206 and Eq. (2) for lead-207, where ^{238}U and ^{235}U are the present planetary content of the two isotopes whose decay constants are λ and λ' respectively, T is the age of the planet, and subscript p denotes primordial Pb, which existed in the solar nebula prior to the formation of planets. *See* LEAD ISOTOPES (GEOCHEMISTRY).

$$^{206}\mathrm{Pb} = (^{206}\mathrm{Pb})_\mathrm{p} + {}^{238}\mathrm{U}(e^{\lambda T} - 1) \qquad (1)$$

$$^{207}\mathrm{Pb} = (^{207}\mathrm{Pb})_\mathrm{p} + {}^{235}\mathrm{U}(e^{\lambda' T} - 1) \qquad (2)$$

Of the four lead isotopes, ^{208}Pb, ^{207}Pb, ^{206}Pb, and ^{204}Pb, the last is not produced in any known process of natural radioactive decay, and is used as an index isotope. For example, Eq. (1) takes the form of Eq. (3), where $\mu \equiv (^{238}\mathrm{U}/^{204}\mathrm{Pb})_{\mathrm{today}}$, and the subscript m denotes a measured ratio of a representative sample. In the Pb-Pb method, the two radiometric clocks expressed in Eqs. (1) and (2) are combined to yield Eq. (4), where 1/137.9 is the present ratio of ^{235}U/^{238}U, a constant reflecting the nonfractionation of the two isotopes. Primitive Pb, supposedly primordial, was found in the troilite phase of the Cañon Diablo iron meteorite, and the composition ratios of such Pb are used in Eq. (4), in which T is then the only unknown. Assuming the existence of a sample representative of the Earth as a whole, it is possible to calculate the actual age of the Earth merely on the basis of the Pb isotopic composition of such a sample. Usually what is calculated is an approximate age for the Earth, because in view of the extensive global and regional differentiation which the Earth has experienced, the assumption of a representative sample is an oversimplification.

$$\left(\frac{^{206}\mathrm{Pb}}{^{204}\mathrm{Pb}}\right)_\mathrm{m} - \left(\frac{^{206}\mathrm{Pb}}{^{204}\mathrm{Pb}}\right)_\mathrm{p} = \mu(e^{\lambda T} - 1) \qquad (3)$$

$$\frac{(^{207}\mathrm{Pb}/^{204}\mathrm{Pb})_\mathrm{m} - (^{207}\mathrm{Pb}/^{204}\mathrm{Pb})_\mathrm{p}}{(^{206}\mathrm{Pb}/^{204}\mathrm{Pb})_\mathrm{m} - (^{206}\mathrm{Pb}/^{204}\mathrm{Pb})_\mathrm{p}} = \frac{1}{137.9}\left(\frac{e^{\lambda' T} - 1}{e^{\lambda T} - 1}\right) \qquad (4)$$

Single-stage model. A calculation based on Eq. (4) represents a single-stage model where terrestrial Pb is assumed to have evolved (through the decay of U) in a single stage (of duration T) from an initial Pb composition identical to the nebula's primordial Pb. It is thus implicit in the model that the Earth accreted directly, or nearly so, from the solar nebula.

In his pioneering work, C. C. Patterson chose oceanic sediments as the representative sample, to which he applied Eq. (4). Repeating that calculation but using the more newly determined values for decay constants and primordial Pb yields Earth age (T) = 4.43×10^9 years and the ratio ^{238}U/^{204}Pb (μ) = 9.3. This terrestrial μ value is quite high when compared with estimated $\mu < 1$ for the solar nebula and $\mu \sim 0.2$ inferred for the source of the Allende chondrite. If it is assumed that the Earth's average composition is chondritic ($\mu < 1$), then it follows from the model that the high terrestrial μ is the result of chemical fractionation that took place in conjunction with the formation of the planet about 4.43×10^9 years ago.

The mechanism of such fractionation is subject to a wide range of speculations. It could be argued that the high terrestrial μ value resulted from loss of Pb by volatilization, possibly during accretion. This conjecture appears in general accord with other observations inferring depletion of the Earth in the volatiles. Alternatively, the high μ may be restricted to an outer layer, the counterpart of which is, perhaps, a deep interior sulfide region which had scavenged primordial Pb from the accreting materials. Since no terrestrial Pb reflecting low μ (say 1 to 3) has ever been sampled, it follows that convection within the Earth may be ineffective in transporting material from the deep interior. *See* EARTH, CONVECTION IN; EARTH INTERIOR.

If, in contrast to the above, $\mu \sim 9.3$ is assumed to be characteristic of the primitive material prior to its accumulation to form the planet, then the calculated age would apply to the primitive material

and not necessarily to the Earth (which could have accumulated at a much later date). This possibility, however, was generally dismissed by geochronologists because it rested on the improbable premise that either the Earth accumulation and subsequent differentiation did not result in U-Pb fractionation, or chemical fractionation did occur but the resulting spectrum of μ values was averaged on the surface of the planet back to the original μ value. Not only were these implausible scenarios dismissed, but a general belief that the Earth is as old as the meteorite (4.56×10^9 years) prevailed, despite the lack of direct evidence. Subsequently, the Apollo program yielded strong evidence for the axiom of a single age for planets and meteorites: differentiated lunar rocks as old as 4.55×10^9 years were found.

Two-stage model. The axiom of 4.56×10^9 years' age applicable to planets and meteorites requires that the younger single-stage age of the Earth is an artifact resulting from the shortcomings of the model which does not take into account the length of the duration of accretion before the onsetting of U-Pb fractionation, and the multistage differentiation history of the planet.

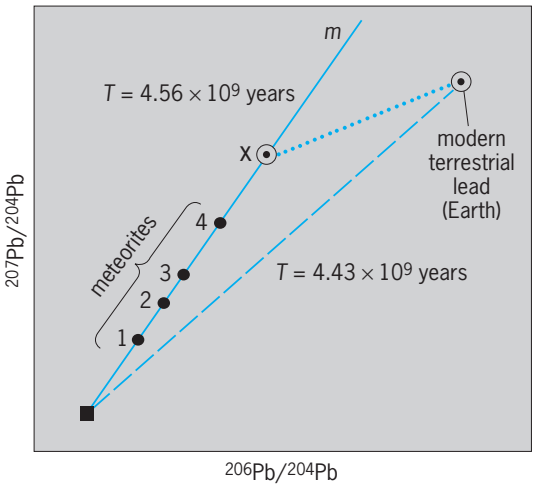
In partial accounting for the complexities, a two-stage model may be applied and may be expressed as Eq. (5), where T_A and T_D are the times of accretion

$$\left(\frac{^{206}\text{Pb}}{^{204}\text{Pb}}\right)_m - \left(\frac{^{206}\text{Pb}}{^{204}\text{Pb}}\right)_p = \mu_1(e^{\lambda T_A} - e^{\lambda T_D}) + \mu_2(e^{\lambda T_D} - 1) \quad (5)$$

and differentiation respectively, and the subscripts 1 and 2 refer to the first and second stage. For simplicity, T_A may be fixed at 4.56×10^9 years. The three unknowns (μ_1, μ_2, and T_D) are solved for in conjunction with knowledge of the ages of uranium-free systems, for example, galenas. The two-stage age of a galena may be given by Eq. (6), where g denotes a galena which was deposited T_g years ago.

$$\left(\frac{^{206}\text{Pb}}{^{204}\text{Pb}}\right)_m - \left(\frac{^{206}\text{Pb}}{^{204}\text{Pb}}\right)_p = \mu_1(e^{\lambda T_A} - e^{\lambda T_D}) + \mu_2(e^{\lambda T_D} - e^{\lambda T_g}) \quad (6)$$

The intricate connection of the parameters is explained with help of the **illustration**. In this schematic presentation, 1, 2, 3, and 4 represent four meteorites falling on a straight line whose slope corresponds to 4.56×10^9 years. The farther away from primordial lead, the higher is the μ of the sample. Modern terrestrial Pb represents the average Pb composition as inferred from surface samples. The slope of the broken line corresponds to the single-stage age of the Earth as given by Eq. (4) According to the two-stage model, modern terrestrial lead did not evolve directly from primordial lead, but rather was shifted from the meteoritic line *m* as a result of U enrichment in a magma reservoir. The time of that differentiation event is given by the slope of the dotted line which intersects line *m* at X. The significance



Schematic diagram of Pb isotopes showing 1–4, which define a line *m*, with a slope corresponding to 4.56×10^9 years. Earth plots off the meteoritic line.

of X is that it gives the present isotopic composition of modern terrestrial lead if the differentiation event did not occur. In other words, point X would yield the correct single-stage age, given by Eq. (7).

$$\left(\frac{^{206}\text{Pb}}{^{204}\text{Pb}}\right)_X - \left(\frac{^{206}\text{Pb}}{^{204}\text{Pb}}\right)_p = \mu_X(e^{\lambda T} - 1) \quad (7)$$

Note that μ_X equals μ_1 of Eq. (5) [as it is the ratio of $^{238}\text{U}/^{204}\text{Pb}$ prior to the differentiation episode], and $T = 4.56 \times 10^9$ years.

The following steps are involved in the calculation: For an assumed value for T_D (that is for an assumed slope for the dotted line in the illustration) the corresponding composition of undifferentiated earth (point X) is obtained from the illustration and the value of μ_X (that is, μ_1) is calculated from Eq. (7). This is entered as μ_1 into Eq. (5) from which μ_2 is then calculated. Finally all three parameters, μ_1, μ_2 and T_D are placed in Eq. (6) to yield a calculated T'_g value. The latter is compared with the known accepted deposition age, T_g, of the galena. The proper parameters are the ones which lead to $T'_g = T_g$.

Using these relationships in conjunction with a large number of conformable galenas and feldspars, the following parameters were calculated for Earth: $T_D = 3.7 \times 10^9$ years, $\mu_1 = 7.2$, and $\mu_2 = 9.7$.

Congruency of conformable galenas. The significance of the two-stage model discussed is in revealing the possibility of a global differentiation episode long after the Earth's accretion. The method of calculation, however, suffers from two main weaknesses: the age of the Earth is assumed rather than determined, and the composition of modern terrestrial lead is not a well-defined entity. In an attempt to overcome these difficulties, the so-called congruency method was put forward. The method provides a simultaneous solution to the ages of cognate galenas as well as the age of the source from which they were derived. Application of the method to conformable galenas shows that they were derived from

two distinct ancient sources, A and B, whose single-stage ages are $T_A = 4.54 \times 10^9$ years and $T_B = 4.43 \times 10^9$ years. Oceanic sediments appear intimately associated with source *B*, which explains the young age obtained from Patterson's calculation.

Sources A and B appear to be counterparts derived from a parental source, denoted AB, through a process involving U depletion in one case ($\mu A = 7$) and its enrichment in the other ($\mu B = 9$) relative to the parental source ($\mu AB = 8$). Significantly, the indicated time of this differentiation is approximately 3.6×10^9 years ago, which appears to substantiate the first-order observation of the T_D calculation. In view of such differentiation, the single-stage ages of sources A and B are gross approximations bracketing the actual age of the primary parental source AB. The age of the latter, as obtained from Eq. (4), is 4.50×10^9 years. This suggests, as an upper limit, an accretion duration of approximately 6×10^7 years prior to the onsetting of planetary differentiation resulting in the formation of the parental source.

Conclusion. There exist about half a dozen modern calculations of the Earth's age based on the Pb isotope systematics. The results fall in the narrow range of $4.43–4.57 \times 10^9$ years, yielding an average age of 4.50×10^9 years $\pm$ 0.07×10^9 years. With a dispersion of a mere $\pm 1.5\%$, the age question seems to be reasonably solved, but active research in this area of geochemistry continues. The primary objective, however, is not the determination of the age with ultimate accuracy, but rather the elucidation of the evolutionary history of the planet. In particular, this type of isotopic research entails the potential of providing strict geochemical constraints which would have to be accommodated by the geophysical models. *See* GEOCHRONOMETRY; METEORITE; RADIOACTIVITY; ROCK AGE DETERMINATION.

Fouad Tera

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Earth, convection in

The notion that thermal convection currents must exist in the Earth's mantle occupies a central position in theories which have been proposed to account for the occurrence of continental drift and sea-floor spreading. Apart from the controlling influence which these currents exert on surface geological processes such as mountain building, they also govern the rate at which heat is transferred from the hot interior of the Earth to its surface and so are crucial to understanding the thermal history of the planet. Mantle convection serves as the guiding principle in the geological and geophysical sciences. *See* CONTINENTS, EVOLUTION OF.

Sea-floor spreading and continental drift. Although even the earliest geographers were impressed by the apparent fit of the coastlines of South America and Africa, it was not until the methods of paleomagnetism had been developed that a plausible case could be made that these two continents had once been joined together as part of the supercontinent Pangea over 100 million years ago. The most important advance in understanding the agency through which breakup of this supercontinent occurred was made in the early 1960s in studies of the pattern of magnetic anomalies observed over a mid-oceanic ridge. The characteristic pattern of magnetic stripes (**Fig. 1**) could be understood in terms of the sea-floor spreading hypothesis, which had been advanced a few years previously. It was suggested that hot material rising from the Earth's mantle beneath the ridge crest acquired an induced magnetization as it was cooled below the Curie point in the Earth's local magnetic field. The material was normally or reversely magnetized, depending upon the polarity of the field at the time of cooling. Although the reason for the episodic reversal of the polarity of the field is not understood in detail, the fact that such reversal occurs allows the sea floor, as it spreads horizontally from a mid-oceanic ridge, to produce a magnetic tape recording of its own large-scale horizontal motion. When the spatial pattern of reversals which is recorded on the sea floor in this way is combined with the time scale for the same reversals established on land, using radiometric dating methods, the pattern of magnetic stripes can be used to measure the horizontal spreading rate as a function of location. When this is done for the globe-encircling pattern of oceanic ridges, local spreading rates are obtained that vary from 0.4 to 4 in. (1 to 10 cm) per year. *See* DATING METHODS; MID-OCEANIC RIDGE; PALEOMAGNETISM; ROCK AGE DETERMINATION; ROCK MAGNETISM.

Paleomagnetic methods have therefore been instrumental in establishing that hot mantle-derived material is continuously rising under and spreading horizontally from the mid-oceanic ridges. That cold and relatively old material is elsewhere descending into the mantle has been demonstrated in an equally convincing fashion by using the methods of seismology. Prior to the advance of the spreading hypothesis, it was suggested that the deep ocean trenches were regions where ocean floor was being downthrust into the mantle along Benioff zones (**Fig. 2**). These zones are defined by the internal locus of deep seismicity. This process is referred to as the subduction of oceanic lithosphere. Taken together with the evidence of spreading, the occurrence of subduction constitutes strong evidence for the existence of a thermally induced circulation in the Earth's mantle. Such a circulation is by definition one which is maintained against dissipative frictional processes by the continuous conversion of potential into

kinetic energy through the action of the buoyancy force. This conversion is effected whenever hot material rises (spreading) and cold material sinks (subduction). The store of potential energy upon which the circulation continuously draws is found in the large increase of temperature between the Earth's surface and the core-mantle boundary. *See* LITHOSPHERE; SUBDUCTION ZONES.

Among the other lines of evidence which have combined to strongly reinforce this model, perhaps the most important concerns the characteristic variations of heat flow and bathymetry across the ocean floor as a function of ocean-floor age. From the many thousand heat flow measurements which have been made in boreholes on the ocean bottom, it is clear that heat flow varies inversely as the square root of the age of the sea floor. Near the mid-oceanic ridges, the heat flow achieves a maximum value which is in excess of 400 mW · m^{-2}, while the flow of heat through the oldest sea floor (age less than 150 × 10^6 years) is between 40 and 45 mW · m^{-2}. These numbers should be compared with the mean heat flow through the continents of about 60 mW · m^{-2}. Ocean-floor depth (bathymetry) varies in a similarly systematic fashion. The shallowest depths are found above the mid-oceanic ridges, and depth increases with the square root of the age of the ocean floor,

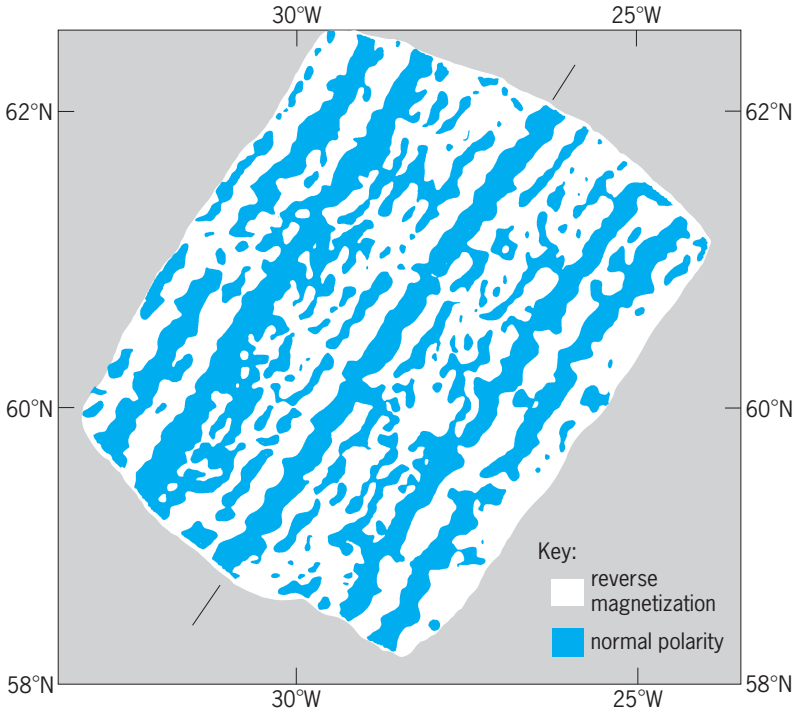

Fig. 1. Pattern of magnetic stripes over the Reykjanes Ridge southwest of Iceland. (*After J. R. Heirtzler, X. LePichon, and J. G. Baron, Magnetic anomalies over the Reykjanes Ridge, Deep Sea Res., 13:427, 1966*)

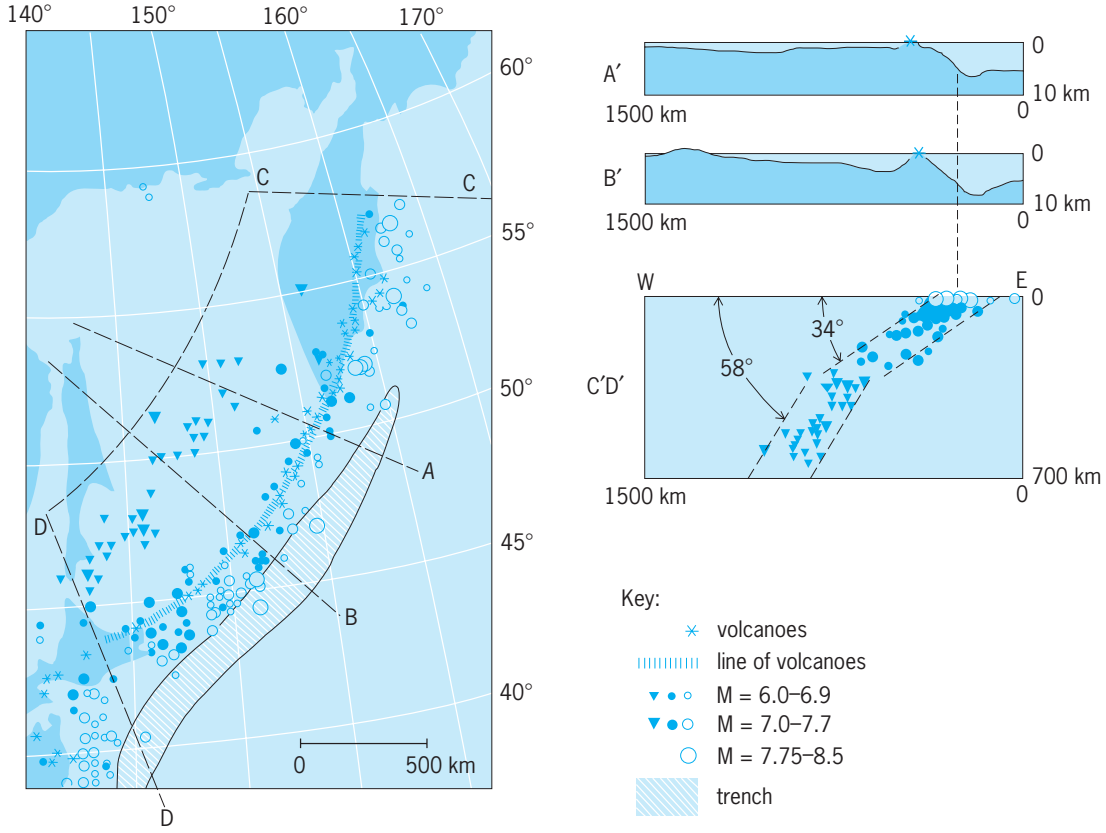

Fig. 2. Geographic setting of the Kurile Trench off the Kamchatka Peninsula, showing the deep ocean trench and the characteristic line of volcanoes in the vicinity of the subduction zone. The sections A′ and B′ illustrate the bathymetry normal to the trench axis, while the section C′D′ shows the localization of earthquake epicenters along the Benioff zone. The surface projections of all earthquake locations are shown in the geographic plan, and their magnitudes *M* are represented symbolically. 1 km = 0.6 mi. (*After H. Benioff, Orogenesis and deep crystal structure: Additional evidence from seismology, Bull. Geol. Soc. Amer., 65:385, 1954*)

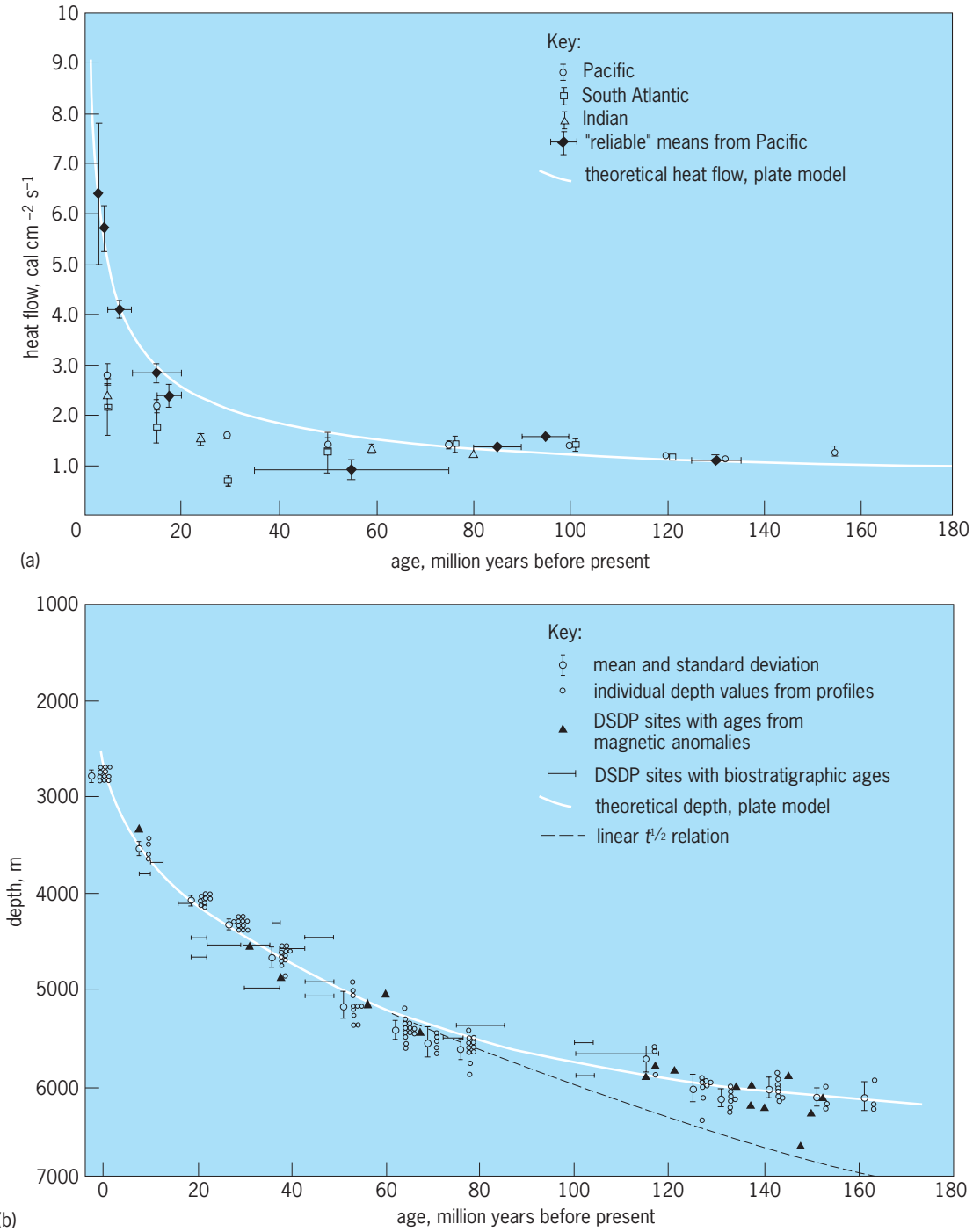

Fig. 3. Characteristic variation of heat flow and bathymetry. (*a*) Variation of surface heat flow as a function of age of the ocean floor in millions of years before present for the Pacific, South Atlantic, and Indian oceans. (*b*) Variation of sea-floor bathymetry with age of the ocean floor for the North Pacific Ocean. For ages less than approximately 80 × 10^6 years, the depth curve follows the simple $t^{1/2}$ decay which is predicted by boundary layer theory for heated-from-below convection. 1 m = 3.3 ft; 1 cal · cm^2 s^{-1} = 42 kW · m^{-2}. (*After B. Parsons and J. G. Sclater, An analysis of the variation of ocean floor heat flow and bathymetry with age, J. Geophys. Res., 82:803, 1977*)

reaching values in excess of approximately 2 mi (3 km) in the oldest oceans (**Fig. 3**). These characteristic variations are precisely the variations to be expected if the ocean floor constitutes an intrinsic part of the cold surface boundary layer of a mantle-wide convective circulation and if this circulation is forced to an important extent by heating from below due to the action of a substantial heat flux across the core-mantle boundary. *See* CONTINENTAL DRIFT; EARTH, HEAT FLOW IN.

Plate tectonics. From a global perspective, attempts to describe the surface manifestations of the underlying mantle convective circulation have given rise to a body of descriptive ideas which have collectively been called plate tectonics. In this kinematic picture, the surface of the Earth is seen to be

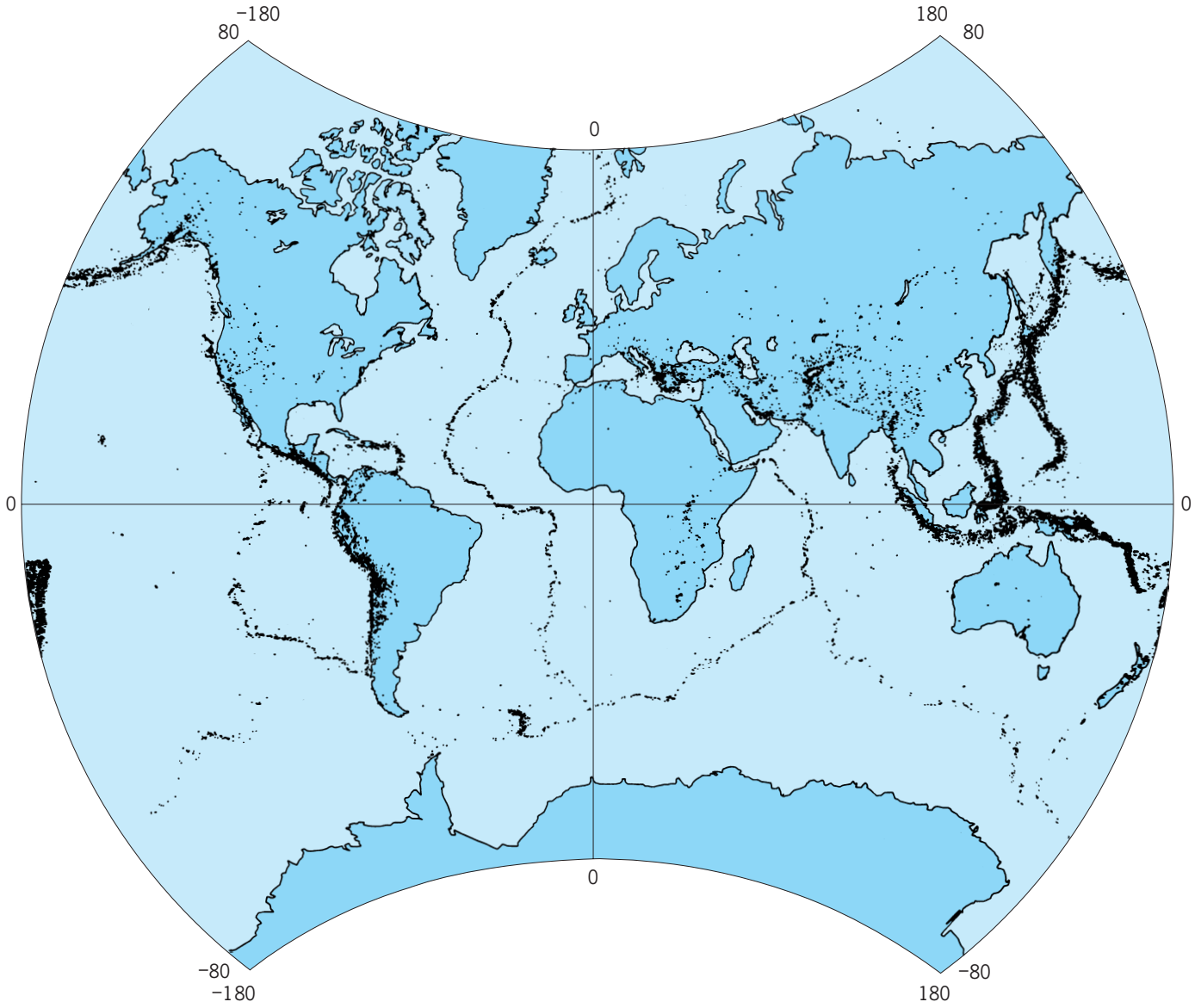

Fig. 4. Global seismicity showing surface projections of all earthquakes which occurred during 1965–1975 with magnitudes in the range 4.5–5.5. Plate boundaries are located along the seams marked by high frequency of earthquake occurrence.

divided into a relatively small number of rigid plates which interact with each other only at their boundaries. Continents are carried passively by the plates and drift with them over the Earth's surface. The boundaries separating adjacent plates are defined on the basis of the global distribution of seismicity, in terms of which it was first recognized that the system of mid-oceanic ridges, deep ocean trenches, and their important interconnections via transform faults was continuous. *See* MARINE GEOLOGY; TRANSFORM FAULT.

Volcanic activity is confined to regions of active subduction along plate margins (**Figs. 4** and **5**). Most major mountain systems are also found in the same regions (the Himalayas, Andes, Alps, and North American Rockies). The idea of plate tectonics is to describe surface geological processes such as mountain building and volcanism as consequences of the interactions among this relatively small number (12) of plates. For example, the currently growing Himalaya Mountains are a product of the collision between the Indian and Eurasian plates. Most volcanic activity is also sharply confined to plate boundaries. The molten rock involved in a volcanic eruption at a mid-ocean ridge is produced by pressure-induced melting when hot material rising in the ascending limb of a convection cell crosses the melting curve as it decompresses. In such locations the lavas are predominantly basaltic. Andesitic volcanoes, on the other hand, are found in the island arcs associated with subduction zones, and their lavas are distinctly different in composition from those found on mid-oceanic ridges. Not all volcanic activity is confined to the plate boundaries. So-called intraplate volcanism, such as that associated with the Hawaiian Islands, also exists, and the explanation of this

Key:

active ridges and continental extensions, minor faults generalized
subduction or overthrust zone, teeth on upper block
major active fault or fault zone, dashed where nature or activity uncertain
rift or zone of normal faulting, teeth on downthrown block
young volcanos (active within last million years, generalized)

Fig. 5. Map of global tectonic and volcanic activity over the past million years. (*After The Physical World, National Geographic Society, Van der Grinten projection, 1975*)

phenomenon remains somewhat enigmatic. One suggestion is that such features are produced by thermal plumes of small horizontal scale, which are formed episodically in the hot lower thermal boundary layer of the circulation in consequence of the temperature dependence of mantle viscosity, and which can rise quickly and adiabatically to the base of the lithosphere, where they induce intense partial melting.

The detailed relation between the surface plates and the convective circulation in the underlying mantle of the Earth is a subject of scientific investigation about which there is no consensus. Most evidence, however, suggests that there is an approximately 1:1 relation between surface plates and deep mantle convection cells, although the three-dimensional geometry of this large-scale flow is expected to be complex. It is perhaps surprising therefore that simple physical models based upon understanding of thermal convection in liquids can provide a general outline of understanding of complex geological processes. *See* PLATE TECTONICS.

Dynamic models of mantle convection. The ability of the mantle convection hypothesis to reconcile the surface observations can be seen most clearly through an understanding of the meaning of the nondimensional parameter called the Rayleigh number, Ra, which determines the strength of the convective circulation which will be caused by a given thermal forcing across a layer of viscous fluid of depth d. This number is defined as shown in Eq. (1),

$$\mathrm{Ra} = \frac{g\alpha d^4 (dT/dz|_b + Qd)}{\kappa^2 \nu} \tag{1}$$

where g is gravitational acceleration acting through the layer, α is the coefficient of thermal expansion, κ is the coefficient of thermal diffusivity, $dT/dz|_b$ is the

temperature gradient at the base of the layer, Q is the rate of internal heating per unit volume (due to radioactive decay processes, for example) divided by the product of density and specific heat capacity, and ν is the kinematic viscosity. Theoretical analysis shows, and laboratory observation confirms, that unless Ra exceeds a critical value Rc, which is normally of order 10^3, thermal convection is impossible.

An estimate of Ra for the whole mantle of the Earth can be made by employing the experimentally determined parameters: $g \approx 4 \times 10^2$ in. $\cdot$ s^{-2} (10^3 cm $\cdot$ s^{-2}), $\alpha \approx 1.4 \times 10^{-5}$ K^{-1}, $\kappa \approx 3.875 \times 10^{-3}$ in.$^2 \cdot$ s^{-1} (2.5×10^{-2} cm$^2 \cdot$ s^{-1}), and $d \approx 1.7 \times 10^3$ mi (2.9×10^3 km) as the mantle depth. The kinematic viscosity ν may be inferred through analysis of postglacial rebound data, which requires a value of ν between 4.65×10^{21} and about 1.55×10^{22} in.$^2 \cdot$ s^{-1} (3×10^{22} and about 10^{23} cm$^2 \cdot$ s^{-1}), with an increase from the low value to the high value across the seismic discontinuity at 402 mi (670 km) depth. Also obtained from the glacial rebound data is a measurement of the thickness $t \approx 60$ mi (100 km) of the layer near the surface, called the lithosphere, in which the effective viscosity is extremely high. The remaining parameters in Eq. (1), $dT/dz|_b$ and Q, are the least well determined, but can be estimated separately on the basis of alternate limiting assumptions. If $Q \approx 0$ in the mantle, an estimate can be made of $dT/dz|_b \approx (T_b - T_s)/d$, where T_b is the temperature at the core mantle boundary and T_s is the surface temperature. Since the outer core is mostly iron, and since it is in a liquid state, a minimum estimate of the temperature T_b may be obtained from the melting point of iron at core-mantle boundary pressures, and estimates are in the neighborhood of 7200°F (4000°C). This could be reduced substantially owing to the presence of a light alloying element, and such a component must exist in the outer core in order to explain the seismically observed mean atomic weight of the material in this region. Using $T_b \approx 3000$°C in Eq. (1) and $Q = 0$, Ra $\approx 10^7$ is obtained. This value is four orders of magnitude larger than the normal critical value of 10^3, so that the mantle is convectively unstable and thermally driven circulations must exist.

The Rayleigh number may also be estimated on the basis of the alternate assumption $dT/dz|_b = 0$, and a lower bound on Q may be deduced by assuming that the heat which leaves the Earth's surface is in equilibrium with the rate at which heat is generated internally by radioactive decay processes. This gives $\rho_m c_p Q = 3q_s/d$, where q_s is the mean surface heat flow of about 80 m W $\cdot$ m^{-2}, and ρ_m and c_p are the mean mantle density and specific heat capacity, respectively. Substituting this value of Q into Eq. (1) yields a value of Ra which is also enormously in excess of the critical value. No matter how the Rayleigh number is estimated, the conclusion that thermal convection currents are expected in the Earth's mantle is inescapable.

The question as to how these convection currents are driven, whether predominantly by heating from below through $dT/dz|_b$ or by internal heating through Q, is not well understood. Also, no consensus exists concerning the depth to which the circulation extends. The heated-from-below model is preferred from several points of view, however, and it will be employed here to demonstrate the ability of the convection hypothesis to reconcile the surface observations described previously. By using boundary layer theory for thermal convection at high Rayleigh number, simple expressions are obtained for the characteristic horizontal velocity of the circulation $\overline{u}$, the thermal boundary layer thickness δ, and the surface heat flux q_s, as shown in Eqs. (2), where the $a_i(\Delta)$ are functions of the aspect ratio of convection Δ, which is the ratio of the horizontal scale w of a convection cell to its vertical scale d. In laboratory convection, the realized aspect ratio is $w/d \approx 1$, and if this is assumed for the mantle circulation with d equal to the mantle thickness, horizontal scales of several thousand kilometers are implied, and in fact, such is the value of the mean horizontal scale of the observed surface plates. If the value Ra $\approx 10^7$ is used in Eqs. (2) and $\Delta = 1$, then $\overline{u} \approx 2.8$ in. (7 cm) per year, $\delta \approx 60$ mi (100 km), and $q_s \approx 100$ mW $\cdot$ m^{-2}. These are very close indeed to the observed surface plate speed, the observed thickness of the lithosphere, and the observed mean surface heat flow. The most important assumption in this simple picture of the convection circulation in the mantle is that the oceanic lithosphere is an intrinsic part of the cold thermal boundary layer of the flow responsible for plate creation and destruction. When this assumption is coupled with the notion that convection fills the entire mantle, the observed mean properties of the circulation are related as they would be in a laboratory convection apparatus. *See* CONVECTION (HEAT).

$$\bar{u} = a_1(\Delta)\mathrm{Ra}^{2/3}\kappa/d \tag{2a}$$

$$\delta = a_2(\Delta)d(Rc/\mathrm{Ra})^{1/3} \tag{2b}$$

$$q_s = a_3(\Delta)(\mathrm{Ra}/Rc)^{1/3}\rho c_p \kappa \Delta T/d \tag{2c}$$

Mantle convection with phase transitions. In order to develop a more detailed understanding of the mantle convention process, beyond that embodied in the simple scaling relations discussed above, it is necessary to construct explicit hydrodynamic models. Most such models are based upon the conventional conservation laws for mass, momentum, and internal energy and the equation of state for a viscous fluid. The equation of state considers the incremental variations of density that exist in the mantle in consequence of any transitions of phase that may be induced by the hydrostatic compression of its constituent material. In fact, the existence of such phase transformations is well established. The most important of these from the perspective of the dynamics of the mantle convection process are the exothermic transition from olivine to the α-spinel structure that occurs near a depth of 240 mi (400 km) and the endothermic transition from β-spinel to a mixture of perovskite and magnesiowüstite that occurs near 400 mi (660 km) depth.

A series of contour maps showing the temperature field in the mantle can be developed from a modern hydrodynamic model of the mantle convection process. The individual images show the mantle in half-section (spanning a time period of about 600 million years) because the numerical model is based upon the assumption that the flow, although occurring in a three-dimensional spherical shell, may be treated as though it were independent of the longitudinal coordinate. The sequence of images demonstrates that the influence of the phase transitions, especially the endothermic transition at 400 mi (660 km) depth, on the convection process is profound. Expressed in the simplest way, the impact of this transition is such as to quite strongly inhibit the flow of mass across this horizon. The degree of inhibition is in fact so strong that the flow spends extensive periods of time in a state in which the circulation is almost perfectly layered. This state is disrupted by episodic instabilities of the thermal boundary layer that develops across the 660-km endothermic horizon when the flow is in the layered state. These disruptions take the form of intense "avalanches'' of cold material from the upper mantle into the lower mantle, during which the vigor of the convective circulation increases significantly. In the axisymmetric model, these episodes of "whole mantle'' convection appear to last for times of order 100–200 million years, while the intervening periods of layered convection last two to three times longer.

If such processes occur in the real Earth, they could be involved in a variety of different and heretofore unexplained phenomena. Principal among these is the so-called supercontinent cycle of ocean basin opening and closing. This cycle and the cycle of orogeny (mountain building) that accompanies it is well expressed in the histogram of occurrence of surface continental rocks as a function of their age. When an avalanche occurs across the 660-km transition, the intense downwelling that occurs at this location causes even extremely distant material on the Earth's surface to collect in the basin above it. This is a plausible mechanism whereby continental fragments may be collected together to form a supercontinent.

Seismic-tomography-based models. In attempting to verify the predictions of the a priori models of mantle convection, discussed above, images of the three-dimensional internal structure of the mantle density field derived on the basis of the application of seismic tomography are important. These methods, which are very similar to the CAT scan techniques employed in medicine, are able to deliver models of the internal variation of a seismic body wave speed, typically that for shear waves. On the basis of the assumption that variations in shear-wave speed may be attributed to variations of temperature alone, the three-dimensional images of shear-wave speed may be mapped directly into variations of temperature. *See* COMPUTERIZED TOMOGRAPHY.

Mathematical expressions developed from a model based on seismic tomographic imaging may be solved to predict a variety of observable characteristics of the mantle convective circulation. The most important characteristics undoubtedly include the planet's gravitational field, the velocities of the surface plates, the topography on the core-mantle boundary, the dynamic component of the topography on the Earth's surface, and the advected component of the radial heat flow. Two such comparisons can be illustrated—the first for the nonhydrostatic geoid and the second for the dynamic topography of the Earth's outer solid surface. Since the existing models of seismic tomography provide information only on the longest horizontal wavelength constituents in the internal density field, and are similarly constrained in their ability to resolve radial structure, the theory can be tested only on these largest scales.

W. R. Peltier

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Earth, gravity field of

The field of gravitational attraction of the Earth. Since, at the Earth's surface, the small centrifugal force due to the Earth's rotation is inseparably superimposed on the attraction, the gravity field is usually understood to include also the effect of the centrifugal force.

Concepts

The gravitational potential V at a point P is the potential energy, due to the Earth's gravitational attraction, of a unit mass situated at P. In other words, V is equal to the work done if a unit mass is brought from infinity to the point P under the influence of the Earth's gravitational field. A mathematical expression is obtained by integrating over the Earth the formula for the potential of a point mass, but this expression is almost useless for practical application. For the potential outside the Earth, an expansion into an infinite series (of spherical harmonics, discussed later in this article) is useful; the principal term of this series is GM/r. This term, in which G is the newtonian gravitational constant (6.67×10^{-8} m^3 kg^{-1} s^{-2}), M is the total mass of the Earth, and r is the distance of P from the Earth's center of mass, represents formally the attraction of a spherically symmetric Earth; the consideration of this term only is not, in general, sufficient. The gravitational force $\mathbf{F}$ on a unit mass is the gradient vector of V, $\mathbf{F} = \text{grad } V$; that is, the components of $\mathbf{F}$ are $(\partial V/\partial x, \partial V/\partial y, \partial V/\partial z)$.

Gravitation and gravity. The resultant of gravitation (pure attraction) and centrifugal force is called gravity. Gravity is the force that acts on the body at rest with respect to the Earth since the effects of attraction and of centrifugal force cannot be separated because of the equivalence of gravitational and inertial mass; thus gravity determines the weight of a body.

The gravity potential W is the sum of the gravitational potential V and the potential of centrifugal force, which is given by a simple analytical expression and may be considered as known.

A body moving with respect to the Earth is also affected by the Coriolis force. Like centrifugal force, the Coriolis force is an inertial force due to the Earth's rotation, but unlike centrifugal force, it does not possess a potential and hence cannot be easily incorporated into the gravity field. Therefore Coriolis force is not considered in the context of terrestrial gravitation. This is perfectly adequate since this force is zero for bodies at rest with respect to the Earth, and almost all measuring systems are at rest. *See* CORIOLIS ACCELERATION; GRAVITATION.

Gravity vector. The gravity vector **g** represents the force of gravity on a unit mass. It is the gradient vector of the gravity potential, $\mathbf{g} = \text{grad}\, W$. The magnitude of the gravity vector is the intensity of gravity, or briefly, gravity g. The dimension of g is force per unit mass, or acceleration. The SI unit is $\text{m} \cdot \text{s}^{-2}$. The cgs unit gal ($1\ \text{gal} = 1\ \text{cm} \cdot \text{s}^{-2}$), named after Galileo, is still frequently used, especially the milligal ($1\ \text{mgal} = 10^{-3}\ \text{gal} = 10^{-5}\ \text{m} \cdot \text{s}^{-2}$). Gravity g on the Earth's surface varies from about 978 gals at the Equator to about 983 gals at the poles. The direction of the gravity vector defines the vertical, or plumb line.

Geoid. The surfaces of constant gravity potential, $W = \text{const}$, are called equipotential surfaces or level surfaces. The surface of a quiet lake is part of a level surface. So is the surface of the oceans, after some obvious idealization; the whole level surface so defined is called the geoid. After C. F. Gauss, the geoid is considered as the mathematical surface of the Earth, as opposed to the visible topographical surface. The plumb lines intersect the level surfaces orthogonally (**Fig. 1**); they are not quite straight but very slightly curved.

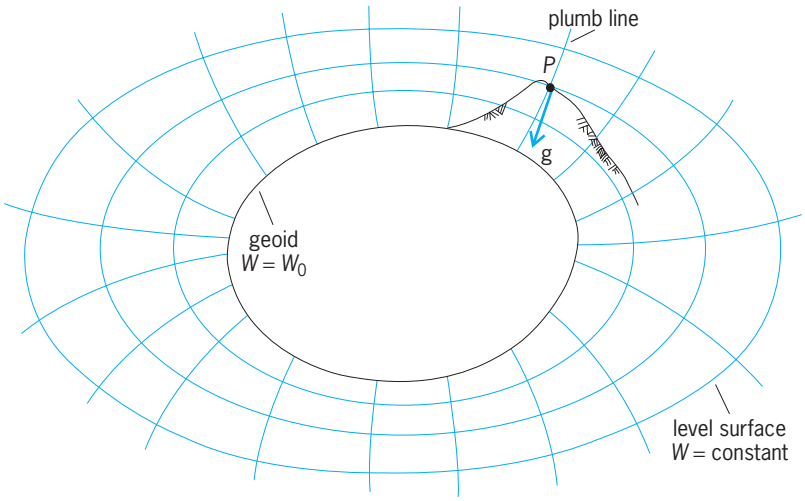

Fig. 1. Level surfaces and plumb lines. The arrow gives the direction of the gravity vector *g* at point *P*. *W* is the gravity potential.

Measurement

The quantity that is measured most commonly is the gravity g. The determination of g as such is called an absolute gravity measurement. Usually only relative gravity measurements are performed, determining the difference between, or the quotient of, the gravity values at two different points. The direction of the gravity vector, which gives the plumb line in space, is measured by astronomical methods. Differences in the gravity potential W are obtained by geodetic leveling. Finally, certain derivatives of g and similar quantities are measured by instruments such as the torsion balance.

Satellite methods have enormously improved knowledge of the gravity field. In the future, new satellite technologies will advance gravity-related measurements, to be studied along with classical theoretical measurements, which will fully retain their importance. *See* GRAVITY.

Classical gravity measurements. Absolute gravity measurements are now exclusively performed by falling-body instruments, which use the fact that the free fall of a test object is proportional to gravity g: if a body, originally at rest, is dropped in vacuum, the vertical distance s covered during time t is represented by Eq. (1), which gives Eq. (2). Since s and t can be

$$s = \frac{1}{2} g t^2 \tag{1}$$

$$g = \frac{2s}{t^2} \tag{2}$$

very accurately measured, gravity g is obtained with a relative accuracy of 10^{-8} to 10^{-9}. Both stationary and movable instruments are used routinely.

The former use of pendulum apparatuses for measuring absolute g gave only a relative accuracy of 10^{-6} and is now completely obsolete.

Relative gravity measurements are performed by spring gravimeters, which are based on Hooke's law: the extension s of the elastic spring is proportional to the force of gravity, so that the difference of gravity g between two points P_1 and P_2 is given by Eq. (3), where k is a factor of proportionality. The

$$g_2 - g_1 = k(s_2 - s_1) \tag{3}$$

same principle is used in accelerometers. The accuracy is consistent with that of modern absolute gravity measurements. *See* ACCELEROMETER; GRAVITY METER.

Since relative gravity measurements are much simpler, absolute determinations of gravity are performed only at relatively few points on the Earth's surface. At other points, gravity is determined relative to these absolute stations. Relative measurements of gravity are made by pendulums or gravimeters.

Basically, an instrument that measures gravity (a gravity sensor) is nothing but a special form of accelerometer. The principle of the spring balance is also utilized for accelerometers, and inversely vibrating-string accelerometers have been proposed as gravity sensors since the frequency of a loaded vibrating string is a function of g. Accelerometers have

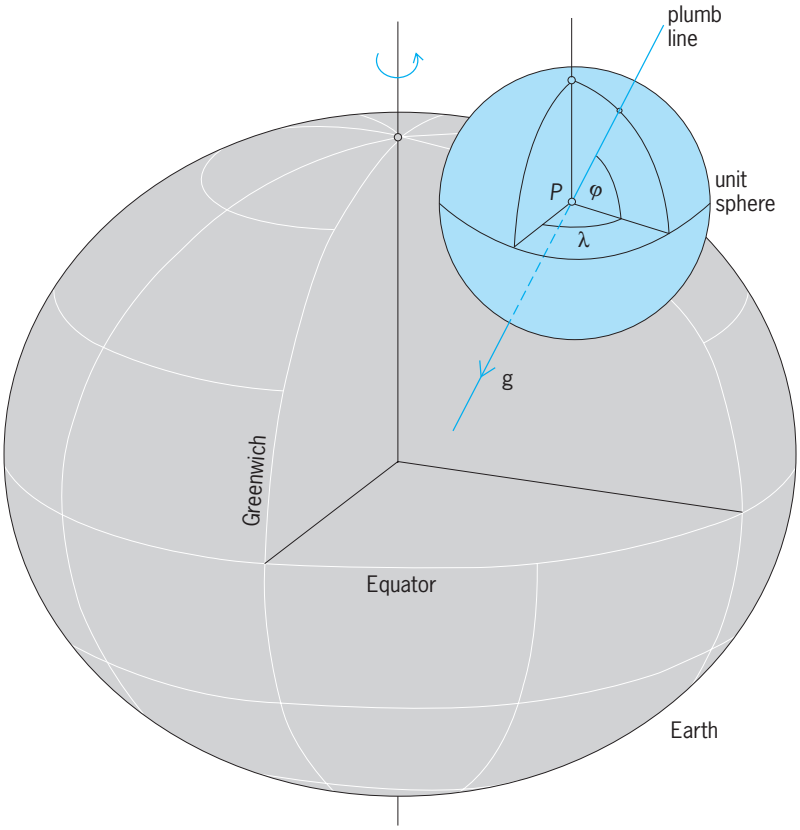

Fig. 2. Direction of plumb line is defined by the geographical coordinates φ and λ. The unit sphere is frequently used for representing spatial directions.

been especially proposed for use in gradiometers, as discussed in the section below on gravity gradients.

Direction of gravity vector. This direction, the plumb line, is defined by two angles; φ, geographical latitude, and λ, geographical longitude (**Fig. 2**). These angles are determined by astronomical methods to an accuracy of 0.1 second of arc or better. If g, φ, and λ are known, then the gravity vector $\mathbf{g}$ is completely determined.

Potential differences. Differences in the potential W can also be measured. If dn is the distance between two neighboring level surfaces of potential difference dW, then $dW = -g\,dn$, so that Eq. (4)

$$W_B - W_A = \int_A^B g\,dn \tag{4}$$

holds. The geodetic operation of leveling (spirit leveling or differential leveling) determines dn. The leveling increments dn are multipled by gravity g and summed along the leveling line that connects points A and B. In this way the difference of potential W between A and B is determined.

Gravity gradients. The gradient of gravity along a certain direction s, $\partial g/\partial s$, is the component of the vector grad g along this direction. Instruments measuring gravity gradients are called gradiometers.

Horizontal gradients (gradients along any horizontal direction) are obtained by means of the torsion balance invented by R. Eötvös about 1900. This instrument consists essentially of a vertically suspended torsion wire carrying a light horizontal beam at which two equal masses are suspended at different heights. The torsion balance gives horizontal gradients of g and, in addition, certain characteristics of the curvature of equipotential surfaces.

The practical application of the torsion balance has suffered from the competition of modern gravimeters, but the interest in gradiometers is being revived because they can be used for gravitational measurements aboard an artificial satellite (a gravimeter would in this case constantly indicate $g = 0$!).

A general gradiometer may be viewed as a combination of two gravity sensors separated by a small distance; the difference in their readings, divided by the distance, gives the gravity gradient. The use of vibrating-string sensors, mentioned above, has been proposed for this purpose. The full power of this technique will be achieved by satellite gradiometry.

Gravitational Field

The gravitational field of a reference ellipsoid of revolution which closely approximates the real Earth is called the normal field. The difference between the real field and the normal field is called the anomalous field.

Normal field. Since the Earth is very nearly an ellipsoid, the gravity field of a suitable ellipsoid of revolution is a good approximation to the gravity field of the Earth and can thus be considered as a normal terrestrial gravity field. The normal gravity potential is denoted by U. An ellipsoid of revolution—the reference ellipsoid—that is close to the geoid is defined to be an equipotential surface of this normal field, where $U = U_0 =$ constant.

Normal gravity. Gravity on such a "level ellipsoid" is called normal gravity γ; it is given by the gravity formula, Eq. (5), where γ_e is normal gravity at the equator, φ is geographical latitude, and ϵ is a very small constant (for numerical values see below).

$$\gamma = \gamma_e(1 + \beta \sin^2 \varphi - \epsilon \sin^2 2\varphi) \tag{5}$$

The other constant, β, is related to the geometric flattening of the terrestrial ellipsoid, $f = (a - b)/a$ (where $a =$ semimajor axis and $b =$ semiminor axis) by A. C. Clairaut's formula, Eq. (6), where $m =$ centrifugal force at Equator/gravity at Equator, in which terms of the second order in f are neglected. This remarkable relation permits the determination of the Earth's flattening f from gravity measurements.

$$f + \beta = \frac{5}{2}m \tag{6}$$

The level ellipsoid and its gravity field are defined by four constants, for example, the set (a, GM, J_2, ω). Here a and GM have been introduced before; J_2 is the dynamical form factor or dynamical ellipticity of the Earth, defined as $J_2 = (C - A)/Ma^2$, where C and A are the Earth's principal moments of inertia; and ω is the Earth's angular spin velocity.

The constants f, β, and γ_e are related to these constants, to the first order in f, by the formulas (7), (8), and (9).

$$f = \frac{3}{2} J_2 + \frac{1}{2} m \tag{7}$$

$$\beta = 2m - \frac{3}{2} J_2 \tag{8}$$

$$\gamma_e = \frac{GM}{a^2}\left(1 + \frac{3}{2} J_2 - m\right) \tag{9}$$

The set of values given below defines the Geodetic Reference System 1980 adopted by the International Union of Geodesy and Geophysics.

$$\begin{aligned} a &= 6,378,137 \text{ m} \\ GM &= 3,986,005 \times 10^8 \text{ m}^3 \cdot \text{s}^{-2} \\ j_2 &= 0.00108263 \\ \omega &= 0.00007292115 \text{ s}^{-1} \end{aligned}$$

The corresponding flattening is $f = 1/298.257$, and the gravity formula becomes Eq. (10).

$$\gamma = 978.0327(1 + 0.0053024 \sin^2 \varphi - 0.0000058 \sin^2 2\varphi) \text{ gal} \tag{10}$$

External and internal field. Gravity above the ellipsoid decreases according to formula (11), where γ_b is

$$\gamma_b = \gamma - (0.30877 - 0.00044 \sin^2 \varphi)b + 0.000072b^2 \tag{11}$$

gravity at elevation b and latitude φ, and γ is the gravity at the same latitude on the surface of the ellipsoid; γ_b and γ are measured in gals, and b is measured in kilometers.

As for the internal field, the internal level surfaces are approximately ellipsoids that become more and more spherical with increasing depth. The theory of the internal field is governed by a differential equation published by Clairaut in 1743. This equation relates the flattening of the level surfaces to the density under the assumption of hydrostatic equilibrium. The solution of this equation offers a possibility of computing the flattening of the Earth from the constant of astronomical precession, which is quite accurately known. Values for f obtained in this way and values obtained more directly from the analysis of satellite orbits, from J_2 [see Eq. (10)], show discrepancies which have yet to be completely explained.

Anomalies of the gravitational field. The normal ellipsoidal field incorporates the main part of the Earth's gravitational field. The difference between the real field and the normal field is called the anomalous field. The anomalous potential T is the difference between the actual gravity potential W and the normal gravity potential U. (The circumstance that the symbol T has been previously used to denote the period of oscillation of a pendulum will cause no confusion.)

Other quantities of the anomalous field are the gravity anomaly and the deflection of the plumb line. Both are more readily accessible to observation than the anomalous potential T, and they may be used to determine T.

Gravity anomalies. The gravity anomaly Δg is defined as the difference between gravity at sea level, g, and normal gravity at the reference ellipsoid, γ: $\Delta g = g - \gamma$. Since measurements are not generally made at sea level, the measured gravity value must be reduced from the Earth's surface to sea level, that is, to the geoid. Depending on the way in which this is achieved, different types of gravity anomalies are obtained.

This reduction may be done without considering the masses below the observation station; that is, g is reduced as if these masses did not exist and the station were "in free air." This is the free-air reduction. The amount of this reduction, to be added to the measured gravity, is by formula (14) approximately $0.309b$ mgals, where b is the elevation in meters.

In the process of gravity reduction, the masses above the geoid may also be removed computationally. This is the Bouguer reduction. The amount to be added to measured gravity is now $0.197b$, the units being the same as before; this figure is based on a rock density of 2.67 g/cm^3.

Instead of completely removing the masses above the geoid, they may be shifted computationally into its interior in such a way as to get a level homogeneous crust according to some theory of isostasy (see later section of this article). This is an isostatic reduction. There are also a number of other gravity reductions.

Deflections of plumb line. The deflection of the plumb line, or deflection of the vertical, is the deviation of the actual plumb line from the normal to the reference ellipsoid. It is characterized by two components, the component ξ in a northern direction and the component η in an eastern direction. If φ_n and λ_n are the normal (ellipsoidal) coordinates, as opposed to their actual counterparts φ and λ, then ξ and η are given by Eqs. (12) and (13). Astronomical observations give φ and λ, whereas φ_n and λ_n may be obtained from geodetic work such as triangulation. The ξ and η so determined are called astrogeodetic deflections of the vertical.

$$\xi = \varphi - \varphi_n \tag{12}$$

$$\eta = (\lambda - \lambda_n)\cos\varphi \tag{13}$$

Anomalous potential and the geoid. The height of the geoid above the reference ellipsoid, the geoidal height N, is intimately related to the anomalous potential T by Bruns's formula, $T = \gamma N$, where γ is normal gravity. Therefore, the determination of T is equivalent to the determination of the geoid.

A relative determination can be made by the astrogeodetic method. Differences $T_B - T_A$ or $N_B - N_A$ are obtained by integrating astrogeodetic deflections of the vertical along a line connecting the points A and B. By this method it is possible to determine the geoid in a limited region from data (plumb-line deflections) in this region only. The astrogeodetic method

furnishes very detailed and accurate geoidal maps (or maps of the potential T) in areas where good astrogeodetic data are available. However, since this method is only relative, the position of the geoid so obtained is determined only apart from an unknown shift with respect to the Earth's center of mass. The use of the astrogeodetic method is restricted to land areas because the necessary data cannot yet be obtained at sea.

An absolute determination of the anomalous potential or the geoidal height can be achieved from gravity anomalies Δg. This is done by an integral formula due to Stokes (1849). Stokes' formula determines the geoid absolutely; that is, the geoidal heights so obtained refer to an ellipsoid whose center coincides with the Earth's center of mass. It requires the gravity anomalies to be given all over the Earth; this is practicable because gravity can be measured at sea as well as on land areas. Large regions, especially the oceans, have not been satisfactorily surveyed gravimetrically. This brings about certain restrictions on the practical applicability of the gravimetric method; nevertheless it has given important results. **Figure 3** shows part of the "Columbus Geoid," a historical landmark computed at Ohio State University.

Harmonic analysis. A periodic function can be expanded into a Fourier series; this is called harmonic analysis. Harmonic analysis may be generalized to functions of more than one variable, such as T or Δg. Instead of Fourier series there is, in this particular case, series of so-called spherical harmonics; such series are applied also in other geosciences, such as geomagnetism.

A series of spherical harmonics is infinite. In practice, there is usually a restriction to series truncated after terms of lower degree. Higher degree terms cannot be determined properly and must be neglected. In this way the general broad features of the expanded function are retained, although fine details are lost.

The spherical-harmonic expansion of T is readily obtained if such an expansion of Δg is available. The practical use of this method is again impaired by the lack of gravity data on parts of the Earth's surface, but this drawback can be largely compensated by a combination of terrestrial and satellite data.

The motion of an artificial satellite in the Earth's gravitational field is most strongly affected by the harmonics of lower degree. Therefore satellite observation gives values of the lower degree spherical harmonics of the potential, which provide a good qualitative picture of the general behavior of the terrestrial gravity field and, in particular, of the large features of the geoid (**Fig. 4**). *See* SPHERICAL HARMONICS.

The practical use of spherical-harmonic expansions for representing the Earth's gravity field is usually expressed in terms of the geoid, and is of basic importance. The expansions are usually obtained by a combination of satellite and gravity data. Satellite orbital data give realistic expansions up to degree of about $N = 30$, with a geoidal accuracy of about ± 5 m (16 ft) and a resolution which can be estimated as follows. One degree of arc corresponds to roughly 100 km (61 mi) of distance of the Earth's surface. A spherical-harmonic expansion to degree of $N = 30$, Eq. (14) gives about $180^\circ/30 = 6^\circ = 600$ km (37 mi)

$$\text{Resolution} \equiv \text{half wavelength} \mathrel{\hat{=}} \frac{180^\circ}{N} \mathrel{\hat{=}} \frac{20{,}000 \text{ km}}{N} \text{ or } \frac{12{,}200 \text{ mi}}{N} \quad (14)$$

which, together with a rms error of about $\delta N = \pm 5$ m (16 ft) is not very satisfactory. This was the state of the art at the beginning of the satellite era, around 1965.

Nowadays, earth gravity models (EGM) combine satellite data with terrestrial gravity and are computed to a spherical-harmonic degree of $N = 360$. This improves the resolution by Eq. (14) to $0.5^\circ = 50$ km (30 mi) [theoretically], the accuracy being about 1 m (3.3 ft) or better. Current models even use $N = 720$ with about the same real accuracy since the same data are used.

Even $N = 1000$ is being envisaged, but without significantly new kinds of measurements it is difficult to really improve the situation. Such new techniques are satellite-to-satellite tracking (SST) and satellite gravity gradiometry (SGG), combined with the Global Positioning System (GPS). These techniques will be outlined below (satellite altimetry has

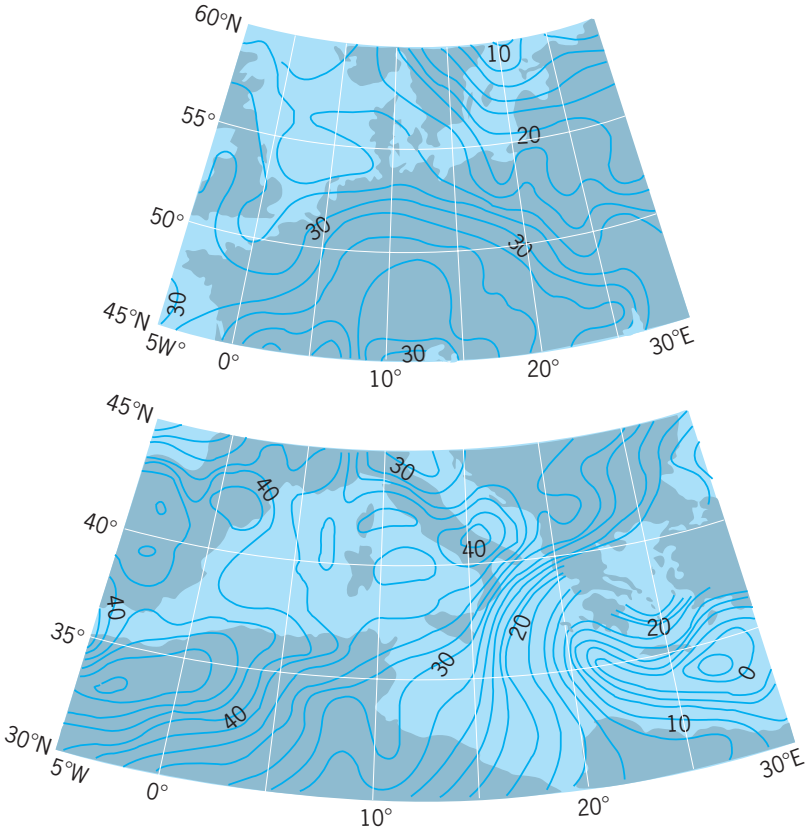

Fig. 3. Gravimetric geoid of Europe, 1957. Geoidal heights in meters (1 m = 3.3 ft), referred to an ellipsoid of flattening 1/297.

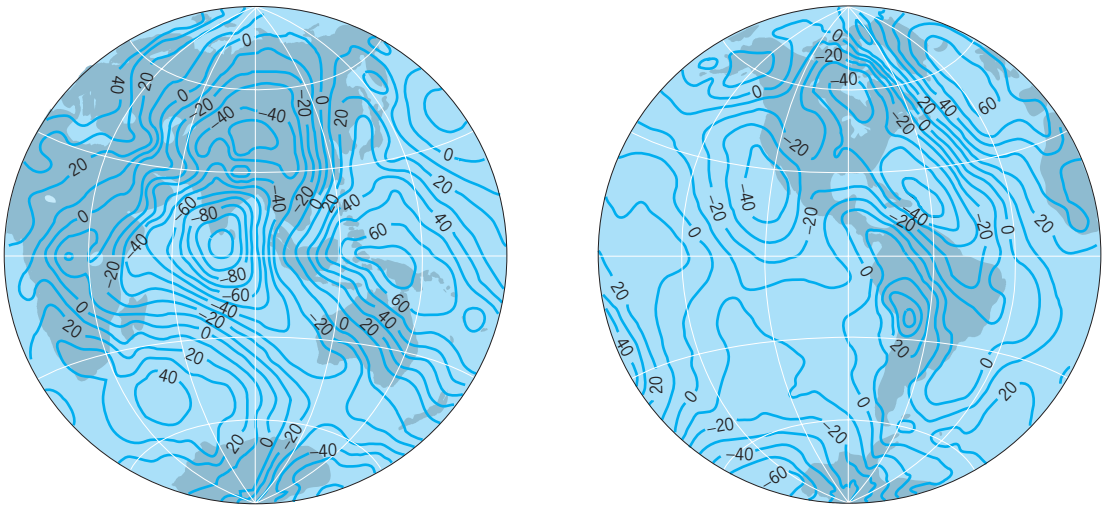

Fig. 4. Geoid from satellite observations combined with terrestrial gravity measurements (EGM96–360). Geoidal heights are in meters (1 m = 3.3 ft), referred to the Geodetic Reference System 1980.

already been taken into account in the $N = 360$ and 720 models).

Satellite altimetry. A direct determination of the oceanic geoid has become possible by measuring the height of a satellite above sea level. This is done by observing the travel time of a radar or laser signal emitted by a satellite and reflected by the ocean surface. If the satellite orbit is known with respect to the reference ellipsoid, the geoidal height follows. A breakthrough in understanding the ocean geoid was achieved by the altimetry satellites *GEOS 3*, 1975–1977, and *Seasat*, 1978.

Satellite-to-satellite tracking. The distance between two satellites changes with location in space due to the irregularities of the gravity field. If continuously and accurately monitored, information of the irregular field may be obtained. Roughly, $N = 100$ could be expected. Implementation will come soon after the year 2000.

Gravity gradiometry. This is an application of the principle of terrestrial gradiometry mentioned above to a gradiometer built into a satellite. Achievable resolutions are higher than with satellite-to-satellite tracking: $N = 250$ is being considered feasible. In combination with other satellite and terrestrial techniques, realistic $N = 720$–1000 models with accuracies on the order of a few centimeters appear possible. Practical application is to come slightly after satellite-to-satellite tracking.

Global Positioning System. GPS and its Russian counterpart, GLONASS, are purely geometrical and therefore (almost) independent of the gravity field. By furnishing precise positions, they make possible a separation of gravitational and geometric effects. In particular, GPS gives inertial accelerations (second derivatives of position) and thus separates them from the truly gravitational forces. This permits putting gravimeters and gradiometers into an airplane whose position is monitored by GPS. At least, airborne gravimetry is already a practically applied reality.

Each of these methods—astrogeodetic, GPS, gravimetric, and the different satellite techniques—has its specific merits and shortcomings. In many respects, they complement each other. By suitable combinations of these data, a very precise gravity field, formulated in terms of a "centimeter-geoid," may become a reality in the near future.

Temporal variations. For most purposes, the Earth's gravitational field may be considered invariable in time. However, it is subject to extremely small periodic variations due to tidal effects. These are caused by the attraction of the Sun and Moon. The attraction acts directly by superimposing itself onto the Earth's gravitational attraction; and it acts also indirectly by slightly deforming the Earth and shifting the waters of the oceans, so that the attracting terrestrial masses themselves are modified.

The lunar effect on gravity attains a maximum of 0.20 mgal, and the solar effect, a maximum of 0.09 mgal; both are well within the measuring accuracy of modern gravimeters. The results of stationary gravimeters recording variations of gravity may be used to draw conclusions as to the elastic behavior of the Earth under the influence of tidal stresses. *See* EARTH TIDES.

Relations to Other Research

The determination of Earth's gravitational field depends on other studies, such as the astronomical determination of the plumb line, and has application in the search for mineral deposits and so on.

Geophysics and geology. The anomalies of the terrestrial gravitational field are caused by mass irregularities. These may be the visible irregularities of topography such as mountains; or they may be invisible subsurface density anomalies. This is the reason why it is possible to use gravity measurements for investigating the underground structure of the Earth's crust. Thus analysis of gravity is applied by geophysicists and geologists for studying general features of the crust, and by exploration geophysicists for searching for shallow density irregularities that might indicate the presence of mineral deposits.

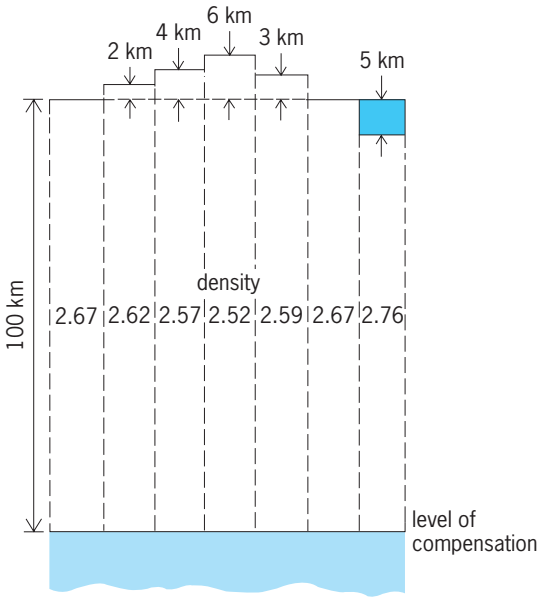

Fig. 5. Pratt isostasy model. 1 km = 0.6 mi.

Isostasy. If the mountains were simply superposed on an essentially homogeneous crust and if the ocean depressions were simply hollowed out from such a crust, the irregularities of the gravity field would be almost 10 times as large as they actually are. This indicates that the visible mass anomalies, such as mountains or ocean depressions, are to a large extent compensated: The crustal density underneath the mountains is smaller than normal, and underneath the oceans, greater than normal.

Theories of isostasy were developed by J. H. Pratt and by G. B. Airy, both about 1850. According to Pratt, the mountains have risen from the underground somewhat like a fermenting dough (**Fig. 5**). According to Airy, the mountains are floating on a fluid lava of higher density, somewhat like an iceberg on water, so that the higher the mountain the deeper it sinks (**Fig. 6**).

Pratt became aware of isostatic compensation through investigating astrogeodetic deflections of the vertical in the Himalayas. At one station he computed from the visible masses a deflection of 28 seconds of arc, whereas the observed value was only 5 seconds of arc.

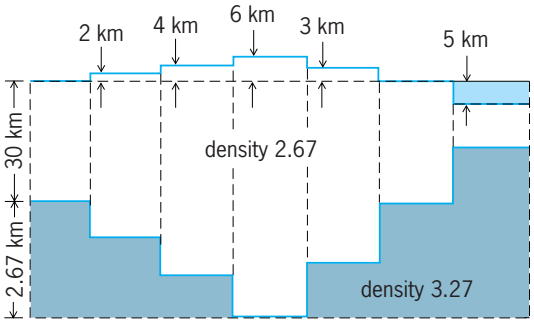

Fig. 6. Airy isostasy model. 1 km = 0.6 mi.

Isostasy is also the reason for a conspicuous behavior of the Bouguer anomalies in mountain areas. Here they are systematically negative and increase in magnitude on the average about 100 mgals/1000 m of mean elevation.

If isostatic compensation were complete, then the isostatic gravity anomalies as defined above would be zero. In general, they are small and fluctuate smoothly around zero; this property makes them particularly well suited for interpolation and extrapolation to poorly observed areas in geodetic applications. Systematically nonzero isostatic anomalies indicate deviation from isostatic equilibrium. Therefore isostatic gravity anomalies, and also Bouguer anomalies, are used to investigate the degree and the mode of isostatic compensation. It appears that the Earth's crust is compensated isostatically to about 90percnt;; the compensation seems in general to be of Airy type rather than Pratt type. *See* EARTH.

Exploration geophysics. Local mass anomalies at shallow depth, which might be related to mineral deposits, are discovered by a local very dense and accurate gravity survey by means of gravimeters. Bouguer anomalies are computed to eliminate as far as possible the effect of visible masses, and various filtering techniques are employed to isolate and localize disturbances. *See* GEOPHYSICAL EXPLORATION.

Geodesy. Geodetic instruments employ spirit levels and other devices to orient them with respect to the horizontal or, what amounts to the same thing, to the plumb line. Since the plumb line is defined by the gravitational field, it can be understood why this field enters essentially into almost all geodetic measurements, even into apparently purely geometric ones. In return, geodetic techniques are among the most efficient means for determining the gravitational field. The "mathematical figure of the Earth" for the purpose of geodesists, the geoid, is defined as a surface of constant potential W. "Heights above the sea level" are heights above the geoid; their determination is therefore a physical as well as a geometric problem. (Geodetic theories have been developed which employ only quantities referred to the Earth's topographical surface; but here the gravitational field enters in an even more complicated way.) Thus geodesy is essentially concerned with the Earth's gravitational field and its determination; the theory of the figure of the Earth is to a large extent equivalent to the theory of terrestrial gravitation. *See* GEODESY.

Astronomy and satellite dynamics. Astronomical observations are indispensable for determining the plumb line, which defines the direction of the gravity vector. Parameters of the Earth's gravitational field such as equatorial gravity γ_e are used to express in kilometers the fundamental astronomical unit, the mean distance of Sun and Earth. The Earth's equatorial radius a, which gives the scale of the geoid, is also used as a basic astronomical constant. *See* ASTRONOMICAL UNIT; EARTH ROTATION AND ORBITAL MOTION.

Terrestrial gravitation affects the motion of the Moon and of the artificial satellites. By observing

these satellites, it is therefore possible to determine essential parameters of the gravitational field, such as the geocentric gravitational constant GM, which is the product of the Newtonian gravitational constant and the Earth's mass. The dynamical ellipticity J_2 defined above, and other lower degree coefficients of the spherical-harmonic expansion of the gravitational potential, may also be determined in this way. GM and J_2 are the coefficients of degrees zero and two of this expansion. *See* SATELLITE (SPACECRAFT).

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Earth, heat flow in

The Earth's heat flow is defined as the amount of heat escaping per unit time from the interior across each unit area of the Earth's solid surface. The movement of heat within the Earth and its eventual loss through the surface are central elements in the modern theory of plate tectonics. The plate movements over the surface of the Earth are seen as one manifestation of a heat engine at work in the interior, and the heat flow at the surface as the exhaust from the engine. *See* EARTH INTERIOR; PLATE TECTONICS.

That the interior of the Earth is a great reservoir of heat is a fact known from antiquity. The eruption of molten rock from volcanoes and the egress of thermal waters and steam from subterranean regions in the form of hot springs and geysers are manifestations of the Earth's internal heat. These phenomena have always been apparent to the human inhabitants on every continent. With the development of underground mining, it became uncomfortably apparent to workers in the deeper levels that the temperature within the Earth generally increased with depth.

Because heat flows from the warmer to the cooler parts of a body, the observation that the Earth's temperature increases with depth clearly implies that heat is escaping from the interior by conduction through the rocky crust. William Thomson (Lord Kelvin) used the heat flow in estimating the age of the Earth. He argued that as the Earth cooled, following its formation and solidification from molten rock, the rate at which the temperature increased with depth (the geothermal gradient) lessened over time, and thus by measuring the geothermal gradient the length of time that the Earth had been cooling could be estimated. Such an estimate is based on the assumption that all the heat being lost is drawn from the endowment of heat that the Earth possessed at the time of its initially molten condition. This assumption was later shown to be untenable because a significant fraction of the Earth's heat flow is derived not from the initial heat but from the decay of radioactive elements present in trace amounts within the Earth. Thus the Earth's heat should be thought of not as a finite quantity acquired at the time of formation and gradually dissipated over time, but as a quantity that in large part has been continually produced, albeit in diminishing amounts throughout the history of the Earth. *See* EARTH, AGE OF; EARTH, CONVECTION IN.

Measurement. The quantitative measure of heat flow, q, is obtained as the product of the geothermal gradient, grad T, multiplied by the thermal conductivity, K, of the material through which the heat is passing: $q = K \text{ grad } T$. The determination of the gradient and conductivity differs in continental and oceanic settings. *See* CONDUCTION (HEAT); HEAT RADIATION; HEAT TRANSFER.

Continents. On continents the geothermal gradient, that is, the rate of increase of temperature downward, is ordinarily obtained by measuring the temperature at several points (or continuously) down a borehole. Most measurements are obtained in mineral exploration boreholes in the depth range 100–500 m (330–1650 ft); at shallower depths there are disturbances to the borehole temperature profile, principally due to circulating ground water, but also arising from the daily and annual temperature variations, irregular distributions of vegetation, topographic relief, erosion and deposition, climatic variations, and other causes. Holes of greater depth, while generally freer of disturbances in the deeper parts, are only infrequently available. Occasionally temperature measurements can be obtained in horizontal borings at different levels of an underground mine, and a geothermal gradient thereby determined. The thermal conductivity is measured on samples of rock obtained from the borehole as it was drilled. The conductivity is usually determined by a laboratory instrument in which a known quantity of heat is conducted through the sample and the temperature gradient established across the sample is measured. Other approaches to the determination of conductivity involve heating the surface of the rock sample with a heating wire or laser for a short time, and monitoring the rise and subsequent fall of the temperature of the rock. *See* DRILLING, GEOTECHNICAL.

Oceans. In oceanic measurements, a hollow probe several meters long equipped with temperature sensors at intervals along the length is allowed to plunge into the soft sediment of the ocean bottom. The temperature of the sediment is thus obtained at several depths, and the geothermal gradient established. As the probe plunges into the soft ocean sediment, a core is collected in the interior cavity and is brought to the surface when the probe is retrieved. The

conductivity of the soft sediment is obtained by inserting a heating needle into it and monitoring the rise of temperature with time as heat is supplied to the core. Some oceanic heat flow probes contain the heating needle internally and allow the direct measurement of conductivity in place while the probe is immersed in the ocean bottom.

The deep ocean floor is a thermally stable environment, generally free from diurnal and seasonal temperature variations that disturb equally shallow temperature measurements on continents. However, it is not free of temperature disturbances arising from the circulation of water through the fractures in the volcanic rock of the oceanic crust. This circulation, actually observed as ocean-bottom hot springs, is most active along the mid-ocean ridges, comprising the youngest areas of the oceanic crust. The hydrothermal circulation on the ridges gives rise in places to the spectacular "black smoker" venting of water at temperatures up to around 350°C (660°F), with important implications for life in the deep oceans. At the ridge crest, unusual biological communities exist at the hot water vents. Because of their relative youthfulness, the ridges have received less of a sediment cover than the older areas of ocean floor, where the ocean water is denied entry into the crustal fractures by a thick sedimentary blanket. *See* MID-OCEANIC RIDGE.

Global data set. In the first review of heat flow measurements in 1954, the total count of measurements both on continents and in oceans numbered only in a few tens. Subsequently, the data set has grown substantially, principally because of the recognition of the role of the Earth's heat in the internal dynamics and surficial tectonics of the planet. Measurements have been made at more than 20,000 sites around the globe in both continental and oceanic settings (**Fig. 1**). On a $5 \times 5^\circ$ grid, the observations cover 62% of the Earth's surface. The most intensive continental coverage has been in the United States, Europe, and Japan, with moderate coverage in Canada, Australia, India, and southern Africa. Gaps include Antarctica and much of Asia, Africa, and South America. The oceanic measurements have been concentrated in the latitude range 60°N to 30°S; the Arctic and southern oceans are only weakly represented with measurements. Despite the uneven distribution and incomplete coverage, a number of useful results bearing on the thermal and tectonic history of the continents and oceans have been obtained from studies of the heat flow data.

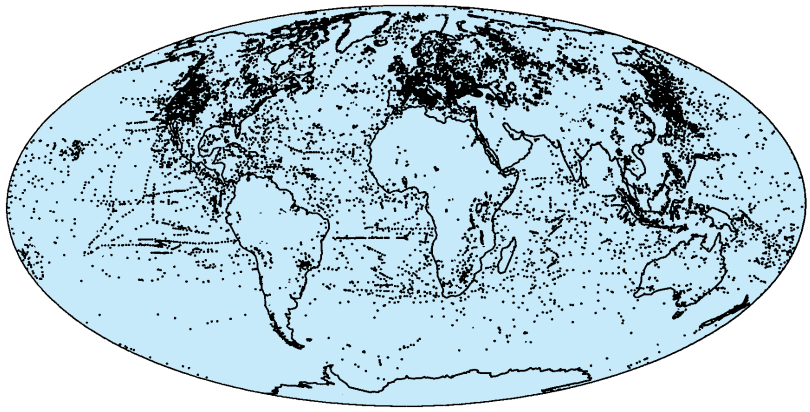

Fig. 1. Geographic distribution of heat flow measurement sites. (*After H. N. Pollack, S. J. Hurter, and J. R. Johnson, Heat flow from the Earth's interior: Analysis of the global data set, Rev. Geophys., 31(3):267–280, 1993*)

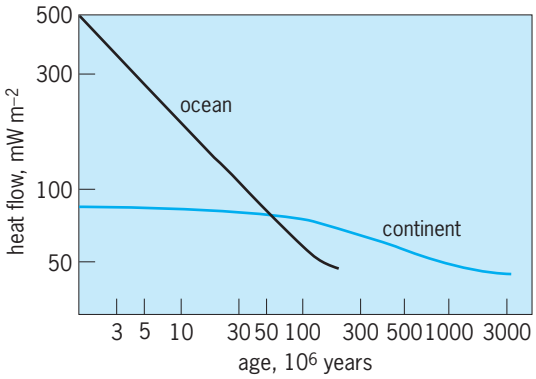

Fig. 2. Heat flow–versus–age relationships for continents and oceans. (*After D. Sprague and H. N. Pollack, Heat flow in the Mesozoic and Cenozoic, Nature, 285:393–395, 1980*)

Heat flow and age relationships. The most significant empirical relationship to emerge from the data studies is that heat flow generally decreases with increasing crustal age at the site of the measurement (**Fig. 2**). In the oceans the significant age is the time of magmatic emplacement of the basaltic crust at a mid-ocean ridge. On the continents it is usually the age of the last major tectonic, magmatic, or thermal metamorphic event to have affected the measurement site. The pattern of the decrease in heat flow with age differs from ocean to continent. While both settings show a heat flow in the older segments of about 40–45 milliwatts per square meter, the oceanic heat flow shows a greater range over a shorter time interval than does the continental heat flow.

The oceanic heat flow–versus–age curve is somewhat obscured in the younger oceanic crust because of seawater circulating through the basalt; however, measurements at sites where sediments have sealed the crust and prevented seawater penetration verify that the inverse square-root dependence on crustal age is valid in the younger crust. The decay of continental heat flow with age extends over a much longer period of time and is more complex than that of the oceans. The continental heat flow principally comprises (1) a radiogenic component derived from the enrichment of the continental crust in heat-producing radioactive isotopes, (2) a component derived from the cooling of tectonothermally mobilized lithosphere, and (3) a background heat flow of probable deep origin. *See* RADIOISOTOPE (GEOCHEMISTRY).

Radiogenic heat in Earth's crust. The heat production from radioactive decay contributes significantly to the heat flow from the Earth's interior. In fact, the entire surface heat flow on continents could arise within the crust if the surface concentrations of the important heat-producing isotopes such as are

Radioisotope concentrations and heat production in the Earth's crust and upper mantle and in chondritic meteorites*

Location	U, ppm	Th, ppm	K, %	Th/U	K/U, $\times 10^4$	Heat production, 10^{-6} W · m^{-3}*
Continental crust	1.25	4.8	1.25	3.8	1.0	0.80
Oceanic crust	0.08	0.20	1.08	2.5	1.3	0.16
Upper mantle	0.02–0.04	0.06–0.15	0.006–0.10	3.0–3.8	0.3–2.5	0.01–0.04
Chondritic meteorites	0.015	0.056	0.09	3.7	6.0	0.025

*U, uranium; Th, thorium; K, potassium; ppm, parts per million.

found in granites and gneisses persisted throughout the full thickness of the crust. However, probable lower crustal rocks such as granulites and migmatites have lesser concentrations of these isotopes, and so radiogenic heat production apparently decreases with depth in some manner. The existence of radiogenic heat and its generation throughout the history of the Earth proved to be the weak link in the Kelvin theory of the age of the Earth.

Radioisotopes. The principal heat-producing isotopes are thorium-232 (^{232}Th), uranium-238 (^{238}U), potassium-40 (^{40}K), and uranium-235 (^{235}U) with respective half-lives of 14.0, 4.47, 1.25, and 0.70×10^9 years. Other radioactive isotopes do not presently contribute significant heat because their decay chains are not sufficiently energetic, or their abundances are insignificant, or their half-lives are too short. The concentrations of uranium and thorium in rocks are generally in trace amounts measured in parts per million, while potassium is much more abundant, in the range of a few percent, of which a small but well-known fraction is ^{40}K (see **table**). A range of values for the upper mantle is characteristic of various possible mantle rocks as inferred from xenoliths brought to the surface by igneous and tectonic processes. If the entire mantle had the lesser concentrations, then crustal and mantle radiogenic heat production would be less than half the present-day heat loss; whereas if the greater concentrations were representative of the mantle as a whole, radiogenic heat production would be about equal to the present-day heat loss.

The abundances of these isotopes are commonly obtained from a multichannel gamma-ray spectrometer, which separates and counts the gamma-ray emissions on the basis of energy. Certain steps in the decay scheme of each isotope involve gamma radiation with characteristic energies, and thus the number of counts at certain levels of the energy spectrum is indicative of the isotopic abundance in the sample. Atomic absorption spectrometry can also be used to determine the amounts of the more abundant potassium, and mass spectrometry to determine the much smaller abundances of uranium and thorium. *See* ATOMIC SPECTROMETRY; MASS SPECTROMETRY.

The explanation for the relatively higher concentrations of uranium, thorium, and potassium in the continental crust resides in the relatively large ionic radii of these isotopes compared to the size of silicon, aluminum, magnesium, calcium, and iron, which combined with oxygen make up the bulk of the Earth's mantle as oxide minerals. The larger ions fit less easily into the compact crystal lattices of the mantle oxides, and in magmatic and metamorphic events they are more easily mobilized and tend to follow the magmatic products and metasomatic volatiles upward into the crust, where they are more easily incorporated in the more open crystal structures characteristic of the lower-pressure environment of the crust.

Heat flow province. The crustal radiogenic component of continental heat flow has been studied extensively. The principal observation is that, within a heat flow province, the regional variation in heat flow is well correlated with the heat production of the surface rocks. The relationship is expressed as $q_0 = q_r + bA_0$, where q_0 is the surface heat flow, A_0 is the heat production of the surface rocks, q_r is the reduced heat flow (the heat flow intercept for zero heat production), and b (the slope of the line) is a quantity with dimension of length that historically has been interpreted as a vertical scale length which characterizes the distribution of the heat sources with depth. The reduced heat flow is then interpreted as that flux that arises below the depth range defined by the empirical scale length. A heat flow province comprises a geographic area in which the heat flow and heat production are linearly related (**Fig. 3**); each heat flow province displays a characteristic reduced heat flow and scale length parameter. However, the traditional one-dimensional (vertical) interpretations have been questioned by investigators who argue that two- and three-dimensional heat transfer can also give rise to an apparent linear heat flow–heat production relationship, depending on both the horizontal and vertical scale lengths of crustal heterogeneities. In the multidimensional interpretations the linear relationship reveals more about the characteristic dimensions of the crustal heterogeneity than it does about the distribution of heat sources with depth.

Global heat loss. The heat flow–versus–age relationships (Fig. 2) that provide the critical constraints for thermal history models of the crust and upper mantle of both oceans and continents can be put to another useful purpose: estimating the heat flow in areas where no measurements have been made but for which the age of the crust is known. In a general way, the distribution of ages in the continental crust has been known for some time, and with the advent of the geomagnetic reversal time scale, large areas

Fig. 3. Geographic distribution of 17 heat flow provinces where the linear relationship between measured heat flow and radiogenic heat production from near-surface rocks has been empirically determined. (*After I. Vitorello and H. N. Pollack, On the variation of continental heat flow with age and the thermal evolution of continents, J. Geophys. Res., 85(B2):983–995, 1980*)

of the oceanic crust have been dated on the basis of their magnetic signatures. Thus it has become possible to estimate the heat flow in unsurveyed areas and map the heat flow over the entire surface of the Earth (**Fig. 4**).

The oceanic ridge system, where new oceanic lithosphere is produced, is important in delivering internal heat to the surface; fully half the Earth's current heat loss comes from oceanic lithosphere produced in the last 66 million years, representing only 31% of the Earth's surface. Other regions of above-average heat flow include the cordillera of western North America, alpine Europe, and the marginal basins of the western Pacific. Areas of low heat flow include all major continental shields and platforms and older ocean basins. Regions where lithospheric plates slide past each other, such as the San Andreas Fault of western North America, show no special heat flow signature, suggesting that friction between the plates is insufficient to generate anomalous heat. The mean heat flow through continents (including marine continental shelves) and oceans is 65 and 101 mW · m^{-2}, respectively. The speculation that the heat flow through oceanic and continental crust is approximately equal, advanced before the relationships between heat flow and tectonic age were well established, is no longer tenable. *See* LITHOSPHERE.

The average heat flow over the entire Earth is 87 mW · m^{-2} and is comparatively a trickle; it is sufficient to bring a thimbleful of water to a boil in about 2 years, or if collected over the area of a football field, would be adequate to light four 100-watt incandescent bulbs. The radiant energy from the Sun intercepted by the Earth is some 4000 times greater than the geothermal flux. The absorption and reradiation of the incident solar energy are the principal processes that determine the surface temperature of the Earth. The geothermal energy is of little significance to the surface temperature but is of paramount importance in considering the Earth's internal thermal condition. *See* GEOLOGIC THERMOMETRY; GEOTHERMAL POWER.

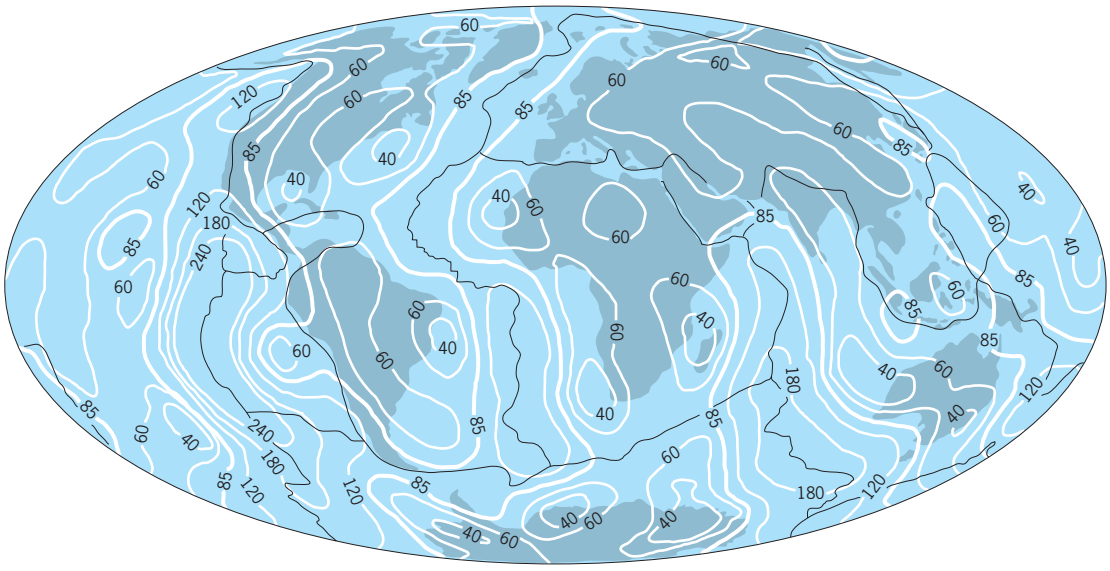

Fig. 4. Present-day global heat flow (degree 12 spherical harmonic representation). Contours are in milliwatts per square meter. (*After H. N. Pollack, S. J. Hurter, and J. R. Johnson, Heat flow from the Earth's interior: Analysis of the global data set, Rev. Geophys., 31(3):267–280, 1993*)

Heat flow in the past. The present-day heat flow from the Earth has commonly been considered a useful if not precise upper bound on the quantity of heat being produced radiogenically in the interior. However, if the heat flow has a temporal variability on a scale that is short compared to the half-lives of the principal heat-producing isotopes, the present-day heat flow must be seen as a much coarser constraint on the Earth's internal heat generation.

Just as the present-day heat loss can be determined from the heat flow–versus–age relationships and the present-day distribution of crustal ages, so can the heat loss at a past time be calculated, provided the age distribution of the crust at that time can be estimated. For the past 180 million years, the age distribution of the oceanic crust at a given time can be determined by subtracting from the present-day distribution all oceanic crust produced from that time to the present and resurrecting and reinstating crust which has been subducted in the interval from that time to the present. The age distribution of the restored lithosphere can be estimated from the present-day age distribution, and the probability that crust produced at an earlier time will have survived to the present day. Continents can be treated similarly, but with different survival probabilities which reflect the longevity of the continents. This method has yielded estimates of the mean heat flow in oceans and continents and the global flux for the past 180 million years. The global mean heat flux, presently 87 $mW \cdot m^{-2}$, has fluctuated in a range of −5% to +15% around the present-day value. The mean heat flow exceeded the present-day flux in the interval 60–100 million years before present, when the rate of production of oceanic crust was high. The mean age of the ocean floor, presently 62 million years, has ranged between 45 and 65 million years, accounting for most of the variation in the global heat loss.

For times earlier than 180 million years ago, an analysis based on the present-day age distribution and survival probabilities becomes unfeasible because in the oceans, where 70% of the global heat loss takes place, there is no oceanic crust surviving with which to estimate past rates of production of the sea floor. Moreover, in studying the Earth's heat loss over intervals of several hundred million to billions of years before the present, it is necessary to consider the half-lives and relative abundances of the principal heat-producing isotopes and the loss of primordial heat acquired during planetary accretion and core segregation. The isotopic endowment of the bulk Earth is unknown but is sometimes assumed to be similar to the chondritic meteorites (see table), which are thought to be representative of the primitive material from which the Earth differentiated. However, the rocks of the Earth's crust have a lower potassium-uranium ratio than do the chondrites. Because of the relatively short half-life of ^{40}K, the heat production 3 billion years ago from a chrondritic composition would have been more than four times the present-day heat production, while a terrestrial composition, based on observed potassium-uranium ratios of crustal rocks, would yield heat at only somewhat more than twice the present rate. *See* METEORITE.

Generalized model calculations of the thermal history of the Earth using the potassium-uranium ratio of terrestrial crustal rocks show that the present-day heat production replaces much but not all of the present-day heat loss. The inequality between heat production and heat loss arises because a finite transit time is required between the time when heat is produced within the body of the Earth and when it escapes across the outer surface, during which interval the heat sources have decayed. The present-day heat loss thus can be seen as older radiogenic heat, delayed in its escape from the Earth's interior by the heat transfer process. Because the replacement heat production is less than the heat loss, the Earth is necessarily cooling. In spite of the various uncertainties about the blend of the nuclear fuel, the thermal history calculations indicate that the heat loss from the Earth 3 billion years ago was at least twice as great as the present-day heat loss. However, the augmented heat flow may not have been distributed uniformly over the Earth. Rather, the increased heat loss may have been best accommodated by a more rapid rate of production of oceanic crust, with only a modest increase of heat flow through the continents. This latter speculation derives from crustal temperatures inferred from metamorphic rocks in ancient continental terrains, which suggest that the heat flow entering the base of the continental lithosphere was approximately the same as today. Henry N. Pollack

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Earth crust

The low-density outermost layer of the Earth above the Mohorovičić discontinuity (the Moho), a global boundary that is defined as the depth in the Earth where the compressional-wave seismic velocity increases rapidly or discontinuously to a value in excess of 4.7 mi/s (7.6 km/s; the upper mantle). The crust is also the cold, upper portion of the Earth's lithosphere, which in terms of plate tectonics is the mobile, outer layer that is underlain by the hot, convecting asthenosphere. *See* ASTHENOSPHERE; LITHOSPHERE; MOHO (MOHOROVICIC DISCONTINUITY); PLATE TECTONICS.

Continental crust. The Earth's continental crust has evolved over the past 4 billion years, and is highly variable in geologic composition and internal structure. The worldwide mean thickness of continental crust is 24 mi (40 km), with a standard deviation of 5.4 mi (9 km; **Fig. 1**). The thinnest continental crust (found in the Afar Triangle, northeast Africa) is about 9 mi (15 km) thick, and the thickest (the Himalayan Mountains in China) is about 47 mi (75 km) thick. Ninety-five percent of all continental crust has a thickness within two standard deviations of the mean thickness, between 13 mi (22 km) and

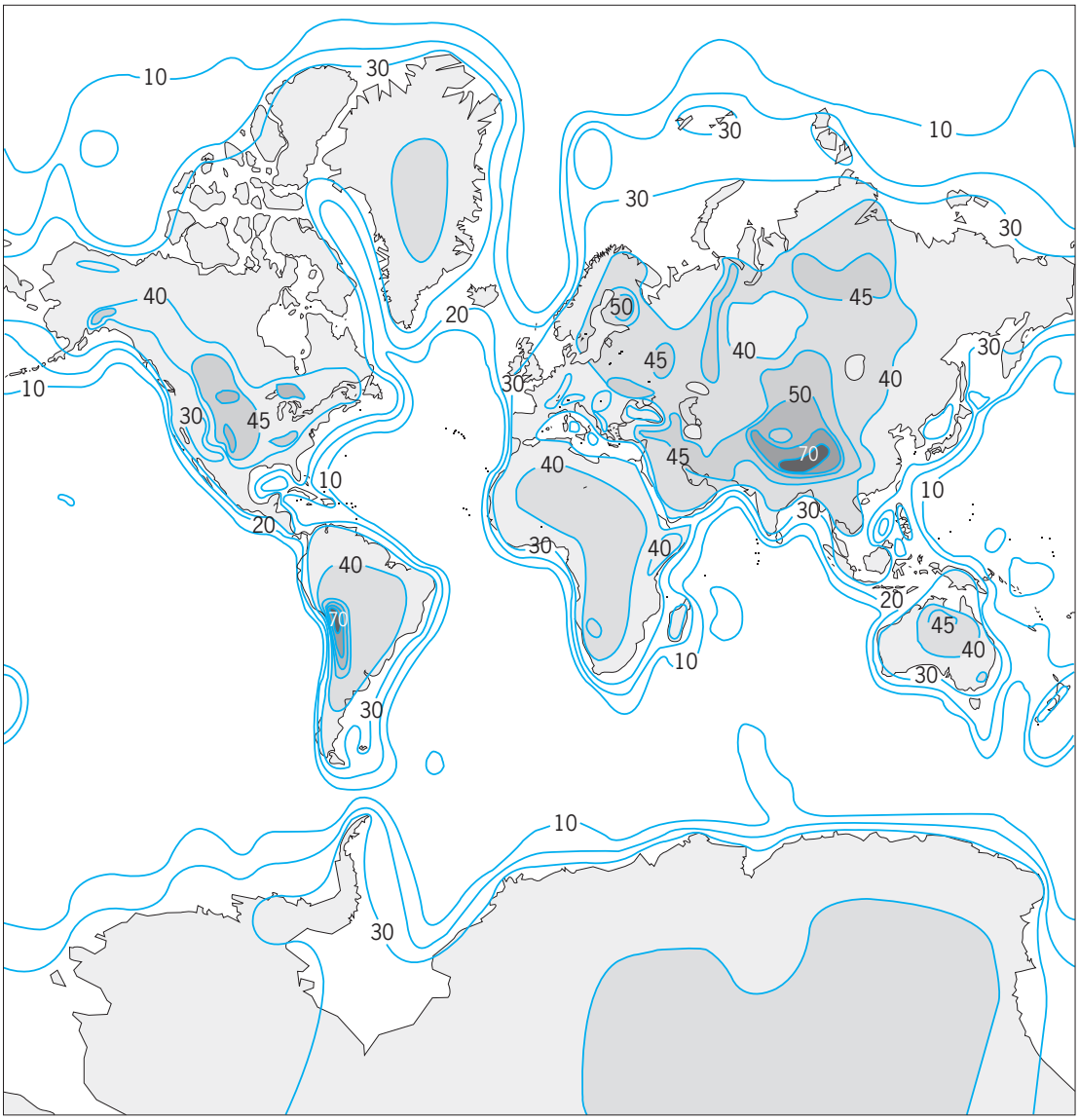

Fig. 1. Mercator projection of crustal thickness (10-km contour interval plus 45-km contour) based on seismic refraction profiles.

37 mi (58 km). The Antarctic has a crustal thickness of 24 mi (40 km) in the ancient, stable (cratonic) region of East Antarctica, and about 12 mi (20 km) in the recently stretched (extended) crust of West Antarctica. Continental margins, which mark the transition from oceanic to continental crust, range in thickness from about 9 mi (15 km) to 18 mi (30 km). *See* CONTINENTAL MARGIN.

Seismic refraction profiles. The internal structure of the crust reveals much about crustal evolution, and one effective way of describing this structure is based on the results of seismic refraction profiles that define crustal layers (that is, layers with characteristic compressional-wave seismic velocity). This seismological method for investigating the crust is widely used for two reasons: seismic refraction measurements are made on a global basis, and the seismically defined layers correspond to geologic layers. There is a relationship between the compressional-wave seismic velocity, density, and rock composition for coarse-grained igneous rocks (**Fig. 2**). Seismic reflection profiles can be used to relate seismic velocities to crustal composition and density. A composite crustal section can be deduced from these data. *See* SEISMOLOGY.

Despite its geologic complexity, the continental crust may generally be divided into four layers: an uppermost sedimentary layer, and an upper, middle, and lower crust composed of crystalline rocks (**Fig. 3**). The sedimentary cover of the continental crust is an important source of natural resources. This cover averages 0.6 mi (1 km) in thickness, and varies from 0 (for example, on shields) to more than 9 mi (15 km) in deep basins. In stable continental crust of average thickness (25 mi or 40 km), the crystalline upper crust is commonly 6–9 mi (10–15 km) thick and has an average composition equivalent to a granite. The middle crust is 3–9 mi (5–15 km) thick and has a composition equivalent to a diorite; and the lower crust is 3–12 mi (5–20 km) thick and has a composition equivalent to a gabbro. Owing to increasing temperature and pressure with depth, the

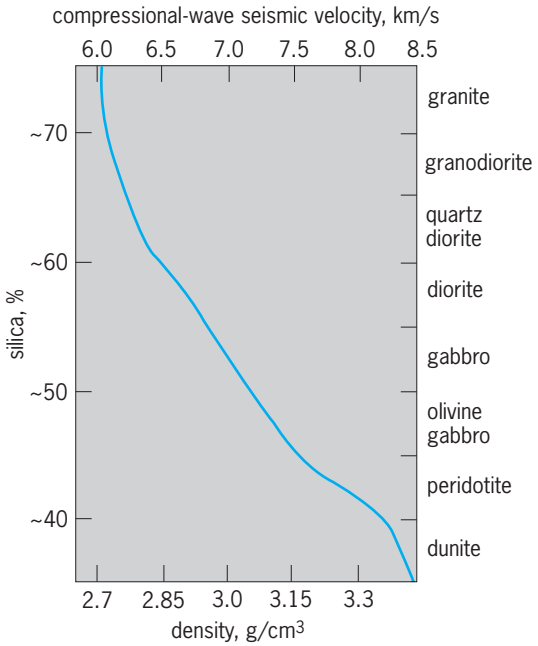

Fig. 2. Relationship between compressional-wave seismic velocity, igneous rock type, silica content, and density. Seismic refraction profiles provide a measurement of compressional-wave velocity as a function of depth in the crust.

metamorphic grade of rocks increases with depth, and the rocks within the deep continental crust generally are metamorphic rocks, even if they originated as sedimentary or igneous rocks (**Fig. 4**). *See* DIORITE; GABBRO; GRANITE; METAMORPHIC ROCKS.

Crustal properties vary systematically with geologic setting, which may be divided into seven groups: orogens (mountain belts), shields and platforms, island arcs (volcanic arcs), continental magmatic arcs, rifts, extended (stretched) crust, and forearcs (Fig. 3). Orogens are typified by thick crust [average thickness is 29 mi (46 km), but the maximum thickness is as much as 47 mi (75 km) in the Himalayas of Asia and Andes of South America]. The seismic structure of orogens is highly variable and is related to the nature of the orogen (for example, continent-continent collision versus ocean-continent collision). However, a thick upper crustal layer is common, and this is the result of thickening of the upper crustal section by wedging and stacking of the compressed upper crust. Shields and platforms, such as the Canadian Shield and the Russian Platform, commonly have an approximately 26-mi-thick (42-km) crust, including a 3–6-mi-thick (5–10-km) high-velocity (4.4–4.7-mi/s or 7.1–7.5-km/s) lower crust. In comparison with shields, island arcs (such as Japan) have thinner crusts and significantly shallower middle and lower crustal layers due to the intrusion of mafic (that is, low silica content) plutons. Continental magmatic arcs, such as the Cascades volcanoes of the northwestern United States, intrude preexisting continental crust, and therefore they are generally 3–9 mi (10–15 km) thicker than island arcs. Continental rifts, such as the East African and Rio Grande rifts, have an average crustal thickness of about 22 mi (36 km), but show great variability in seismic structure. Extended continental crust, such as the Basin and Range Province of the western United States, averages 18 mi (30 km) in thickness, and has relatively little high-velocity (greater than 4.2 mi/s or 6.8 km/s) crustal material. Forearcs are regions that were formed by the coast, near volcanic arcs, including much of the west coast of North America. They typically have thin crust,

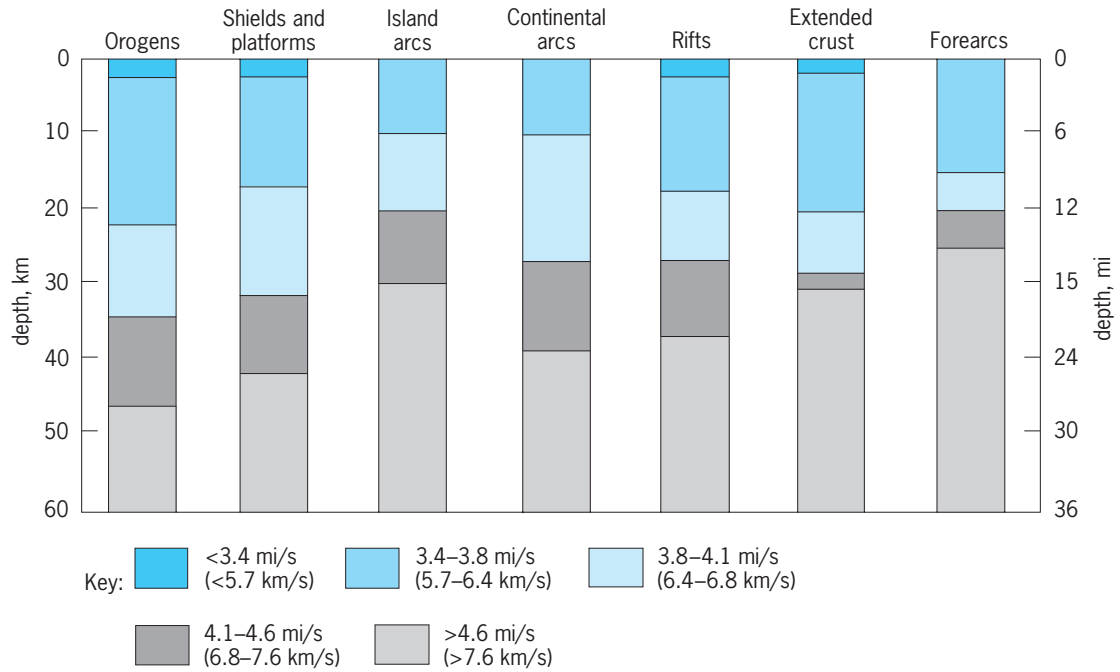

Fig. 3. Average crustal columns for seven geologic provinces, showing variations in crustal thickness (light gray areas correspond to the mantle). Compressional-wave velocity, which has been divided into five ranges, is directly related to geologic composition, as indicated in Fig. 2. (*After W. Mooney et al., JGR, 1998*)

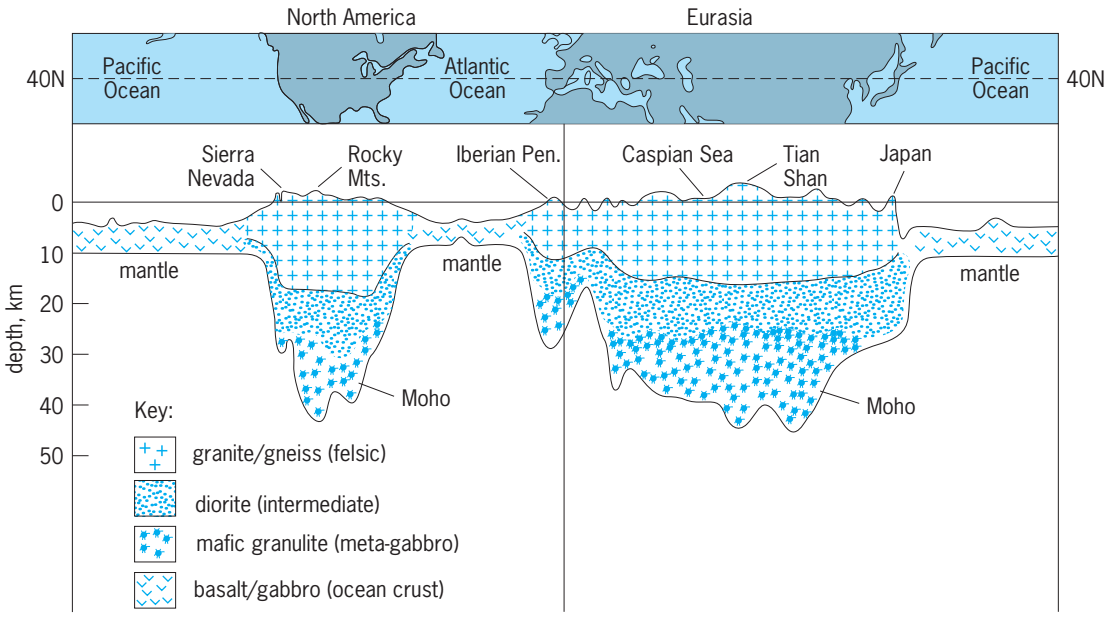

Fig. 4. Cross section of the Earth's crustal thickness at 40°N. Felsic rock is found throughout the upper continental crust above diorite and mafic granulite. The crust is thicker beneath mountain ranges and other regions of high elevation. Vertical thickness relative to horizontal distance is exaggerated 100×. (*After W. Mooney, IASPEI, 2004*)

about 15 mi (25 km), and have a thick (9 mi or 15 km) upper crustal section that consists of relatively low-density metasedimentary rocks. *See* NORTH AMERICA; OCEANIC ISLANDS; PLUTON; RIFT VALLEY; SEDIMENTARY ROCKS; VOLCANO.

Seismic reflection profiles. Seismic reflection profiles use vertically traveling waves to obtain a high-resolution image of the crust, whereas refraction profiles use horizontally traveling waves to obtain seismic velocity information. Reflection profiles provide images of the deep crust that may be interrupted in terms of lithologic, metamorphic, and porosity layering, dipping fault zones, and (more rarely) the presence of crustal melts. Distinct reflection patterns that correlate with geologic setting can also be identified. These patterns provide important information on how the crust has been formed and modified. In many regions, particularly in extended crust, the lower crust contains numerous horizontal reflections that terminate at the depth of the Moho. These reflections are most likely the result of igneous intrusions into the lower crust, combined with horizontal stretching of the warm and ductile lower crust. *See* IGNEOUS ROCKS.

Seismic reflection profiles across mountain belts (orogens) show near-horizontal sheets of rock (nappes) in zones of crustal thickening and deformation. These upper crustal nappes can be followed to midcrustal levels in the Alps, and what has been termed thin-skin tectonics [the horizontal transport of 3–6 mi-thick (5–10-km) sheets of upper crustal material over large distances] has been confirmed in the Appalachians of the eastern United States and in western Europe. *See* CORDILLERAN BELT; MOUNTAIN SYSTEMS.

Seismic reflection profiles in Precambrian crust show pronounced, widespread structural features dating from that period, indicating long-term thermal and mechanical stability. It is possible that the geometry of these structures is qualitatively consistent with plate tectonic processes operating 2 billion years ago. *See* PRECAMBRIAN; SEISMOLOGY.

Geoelectrical measurements. Measurements of electrical conductivity within the deep continental crust have revealed high conductivities at midcrustal depths that provide important information on crustal properties. High crustal conductivity can be caused by several factors, including the presence of saline fluids, partial melts, sulfide minerals, or carbon films. In regions of high heat flow, such as actively extending crust, high crustal conductivities are often found at midcrustal depths (6–12 mi or 10–20 km), and may indicate the melting of wet granitic rocks or the presence of fluids produced by metamorphic dehydration reactions. In young mountain belts, such as the Carpathian Mountains of eastern Europe and the Alaska Range of central Alaska, dipping high-conductivity zones can be correlated with large carbon-rich flysch and black shale basins that have been crushed and deeply buried during the mountain-building process. *See* BLACK SHALE; GEOELECTRICITY; OROGENY; TURBIDITE.

Continental crustal evolution. The properties of the continental crust provide important constraints on the process of crustal evolution. The continents coalesce and disperse as ocean basins open and close, and continental crust is deformed and modified during rifting and at ocean-continent and continent-continent collisions. A net increase in the total volume of continental crust requires extraction of crustal material from the mantle and the stabilization of this material within a continental block.

At least three processes provide new continental crust. The first is the accretion and consolidation of island arcs, such as Japan or the Aleutian Islands, onto a continental margin. The second process is the

tectonic underplating of oceanic crust at active subduction zones. In this process, the continental crust grows from below as oceanic crust is welded to the base of the continental margin, either when subduction stops or when subduction steps oceanward and a new trench is formed. This process has been identified in western Canada and southern Alaska. The third process is the magmatic inflation of the crust at continental arcs, rifts, and regions of crustal extension. This process has been identified in many regions. *See* GEODYNAMICS.

Although both the accretion of island arcs and tectonic underplating of oceanic crust can be documented as stages of continental growth, these processes by themselves would result in a bulk crustal composition that is more mafic (low silica content) than the actual bulk composition of continental crust. This apparent discrepancy can be resolved in at least two ways. First, the mafic lower crust of some island arcs may be delaminated (peeled off from below) and subducted back into the mantle, thereby leaving the remaining crust less mafic. Second, an originally thick mafic lower crust may undergo secondary differentiation that produces an intermediate composition melt and a mafic-ultramafic residue (restite), which ponds at the base of the crust. After cooling, the restite would have a seismic velocity of about 5 mi/s (8 km/s) and therefore would constitute a new, shallower seismic Moho with a less dense, less mafic overlying crust.

Walter D. Mooney

Oceanic crust. The surface of the ocean crust, except for some locally high volcanoes and plateaus, resides some 1–3 mi (2–5 km) below sea level, approximately 3–6 km below the average level of the continents. While the continental crust consists mainly of ancient mountain belts, with an average age of 1500 Ma, oceanic crust is relatively young and dynamic, having a maximum age of only 200 Ma. Most of it was produced at mid-ocean ridges during sea-floor spreading, where ridges define the accret-

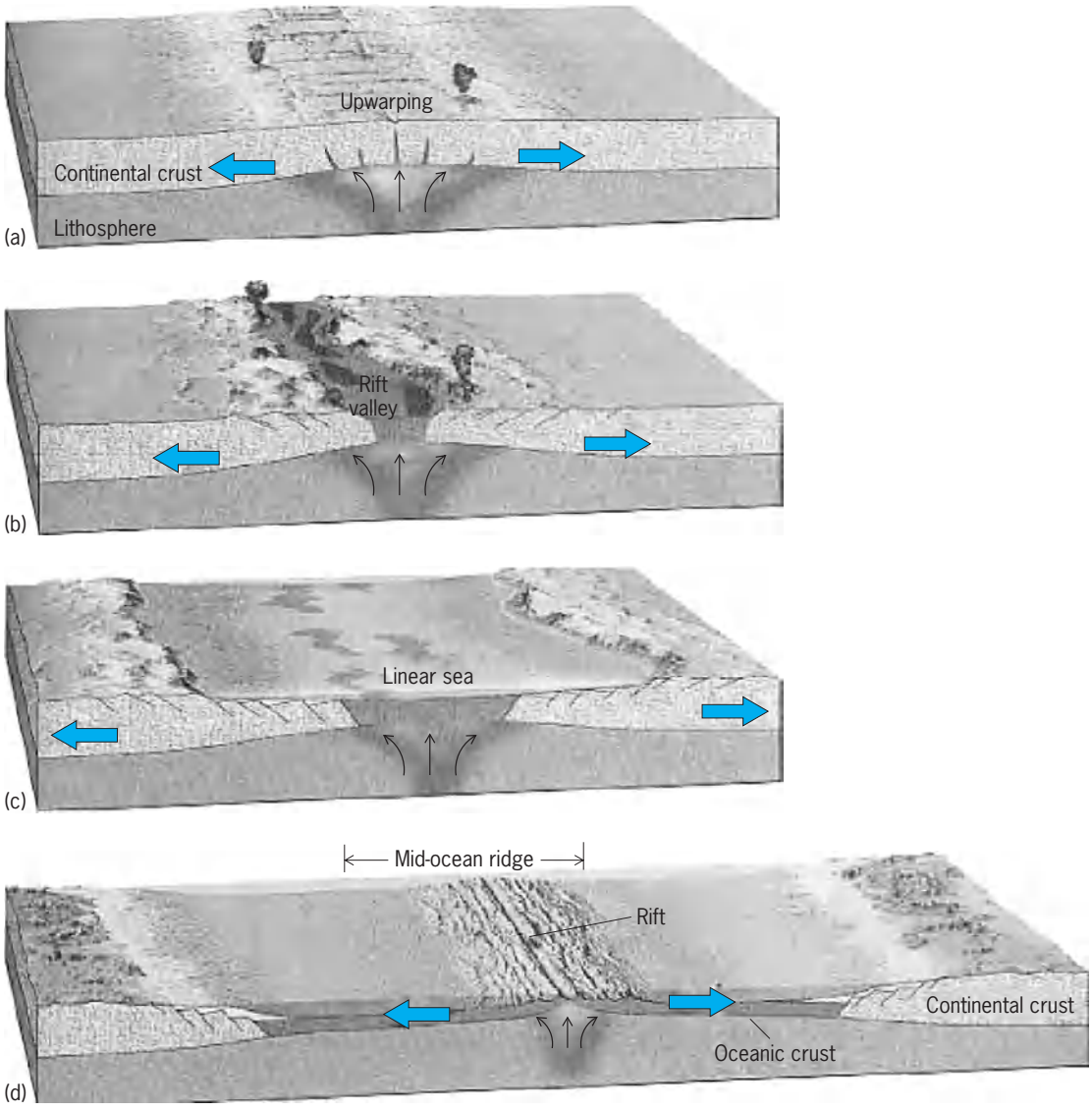

Fig. 5. The sea-floor spreading mechanism begins with continental rifting initiated by magmatic processes (*a*). Once a basin has rifted apart significantly (*b*), just as in the East African Rift Valley, it may fill with water, eventually forming a lake or sea (*c*). Over time, the extrusion of magma forms new crust and the ridge continues to spread, creating a vast ocean (*d*).

ing boundaries of major lithospheric plates moving about the surface of the Earth. Thus, the oldest rocks of the ocean crust date back no earlier than the rifting episodes that initiated the most recent phase of continental drift (the Pangaean breakup) in Late Jurassic times. *See* JURASSIC; MID-OCEANIC RIDGE.

Sea-floor spreading. The sea-floor spreading mechanism holds that the ocean basins, for example the Atlantic and Indian oceans, developed after continental rifting. During sea-floor spreading, magmatic upwelling at the central mid-ocean ridge repeatedly introduces new basaltic crust along a narrow median rift, in the center of a basin within a global system of underwater mountains. This system of mountains, which is over 12,000 mi (20,000 km) long, is interconnected through all the major ocean basins, including the Pacific and Arctic basins. It has been determined that this process of rifting, basaltic magmatism, and continued sea-floor spreading has produced the vast majority of ocean crust (**Fig. 5**). The rates of plate motions are deduced from the records of Earth's magnetic reversals preserved in the basaltic rocks surrounding ocean ridges. While magnetic domains in liquid magma will rotate according to the Earth's magnetic field, once lava solidifies, the orientation of material is no longer flexible and the basalt will "freeze," reflecting the direction of the Earth's magnetic field at its time of emplacement. The anomalies, which are mapped with magnetometers towed from ships, document the shifting patterns of plate motion in the ocean basins, as well as the history of sea-floor spreading. They demonstrate that the ocean crust is one vast, ~6-km-thick (4-mi) basaltic lava field covering nearly 70% of the Earth's surface. *See* BASALT; BASIN; CONTINENTAL DRIFT; EARTH, CONVECTION IN; LAVA; MAGMA; MAGNETOMETER; ROCK MAGNETISM.

The extent of this lava field and its fundamental basaltic composition have been documented by extensive dredge sampling programs along active ridges, the transform faults that offset them, and the fracture-zone scars that trail far off onto the flanks of the ridges. The older portions of the ocean crust have been sampled in dozens of localities by drilling through a carapace of mainly biogenic marine sediments. In all cases, the continuity of the spreading process has been confirmed. Thus, emplacement of basalt at mid-ocean ridges is the universal mechanism by which ocean crust is formed. *See* MARINE SEDIMENTS; TRANSFORM FAULT.

Characteristics. The base of the ocean crust is defined by the location of the seismic Moho. The Moho is usually an abrupt transition between rocks with average seismic velocities of about 4.2 mi/s (6.7 km/s) to rocks with velocities of about 5 mi/s (8 km/s). The former velocities agree with experimental determinations on varieties of nonporous mafic rocks including gabbros and amphibolites (rocks that have significant proportions of plagioclase and little olivine), whereas the latter velocities can represent only nearly fresh (nonserpentinized) peridotite, consisting mainly of olivine and pyroxenes with virtually no plagioclase. In the oceans, explosion seismology and seismic reflection carried out from ships are used to determine both the location of the Moho and the structure of the overlying crust. Both gabbros and serpentinized peridotites are present at various tectonic exposures in the ocean crust, such as along fracture zones, the walls of median rifts, and low-angle detachment surfaces in the floors of rifts. Thus, it is likely that gabbros are the rocks of the lower ocean crust that overlie mantle peridotites. However, there is no known place in the ocean basins where a coherent crust-mantle transition is exposed in outcrop. *See* AMPHIBOLITE; PERIDOTITE.

There are fault slices of former ocean crust on land, known as ophiolites, where nearly complete cross sections through the crust can be mapped and sampled (**Fig. 6**). These are robust indications that the ocean crust consists (from top to bottom) of submarine extrusives (usually pillow basalts), feeder dikes (often vertically sheeted), or sills, gabbros, and peridotites. There is much uncertainty, however, about the extent to which typical ophiolites, most of which formed in island-arc or backarc environments, can represent abyssal ocean crust, which is produced at the major accreting plate boundaries. Moreover, due to the extent of alteration and metamorphism, particularly in ultramafic sections, the rocks in ophiolites are sometimes not well correlated with those in ocean crust. Thus, seismic models of the ocean crust and the nature of the Moho transition therein cannot be determined directly. *See* OPHIOLITE.

Nevertheless, the general seismic model of the ocean crust and the basic ophiolite model of crustal stratigraphy have generally been confirmed both by direct observation and by drilling. Sheeted dikes, for example, have been observed from submersibles on

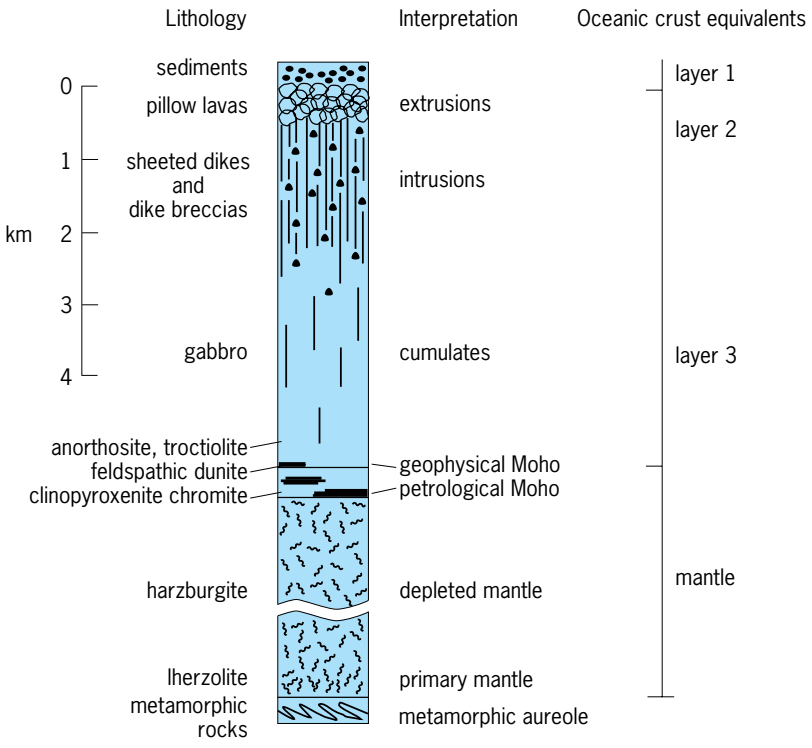

Fig. 6. Interpretation of Bay of Islands ophiolite crust as oceanic crust and mantle, with the various thicknesses estimated from geologic mapping. (*After R. G. Coleman, Ophiolites, Springer, 1977*).

faulted exposures in both the equatorial Atlantic and eastern Pacific. Seismic velocities in the upper part of the ocean crust are consistent with the presence of pillow lavas or flows with many cracks and cavities. Velocities increase systematically with depth, indicating that this porosity structure diminishes, primarily because of compaction, filling of veins with alteration minerals, and transition to compact dikes. The deepest hole drilled in ocean crust, on the flank of the Costa Rica Rift in the eastern equatorial Pacific, reflects these relationships in downhole logging measurements through 1 mi (1.8 km) of igneous material. The transition to rocks with velocities of about 4.2 mi/s (6.7 km/s) is actually within the compacted dike sequence, and does not formally correspond to the lithologic transition to gabbros, which still has not been identified. *See* PETROLOGY; STRATIGRAPHY.

Crustal formation. The composite lithology of gabbros, dikes, and extrusives forms ocean crust above the mantle in the following way. Gabbros comprise the frozen walls, roof, and floor of a spreading ridge's magma-chamber complex, which is more or less continuously supplied with magma from the upper mantle. The flux of magma, and thus the details of this process, differ according to whether the ridge is spreading rapidly or slowly, or is near an unusually hot region of mantle (that is, Iceland). The magma is supplied because of decompression melting of hot peridotite as it convects upward or rises diapirically into the region beneath the spreading ridge. Melts are driven out of the mantle rocks because they are under stress. While they are still molten, they coalesce, flow into fissures, and feed into the lower crust, where they cool and begin to crystallize. *See* DIAPIR; HOT SPOTS (GEOLOGY).

The steady-state process of sea-floor spreading, however, requires that magmas be supplied to the lower crust nearly continuously. This forces magmas already residing in the crust out toward the surface. Elastic materials will initially respond to stress by bending or stretching, but not by breaking (similar to taffy or putty), in a process called ductile deformation. If the rate of stress is fast enough, then rocks will fracture under brittle deformation. At high temperatures (perhaps 1100–1300°F or 600–700°C), rocks will tend to experience only ductile deformation. At lower temperatures, they will more likely experience brittle failure. Earthquakes, in particular, are evidence for large-scale brittle failure at some depth in the crust or upper mantle.

Rocks in the uppermost crust, chilled by hydrothermal circulation of seawater, often experience brittle failure along vertical fractures under the combined impetus of upward magma injection and sea-floor spreading. The tensional stress regime at spreading ridges determines that most fracturing will occur in a narrow zone parallel to the ridge (**Fig. 7**). Thus, dikes tend to be sheeted, and when these breach the sea floor, eruptions occur. Lava flows through dikes will be interrupted by more dikes as the ridge continues to spread, and repeatedly buried by younger flows after being carried to one side of the ridge or the other. Observations made from drilling and submersible instruments indicate

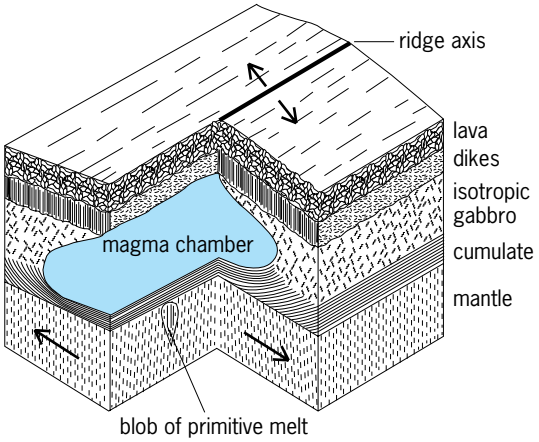

Fig. 7. Open-system magma chamber that may exist beneath all active mid-oceanic ridges, except those that are spreading extremely slowly. (*After M. Wilson, Igneous Petrogenesis, Routledge, Chapman, and Hall, 1988*).

that this process takes place perhaps 10–20 times in the course of building a given segment of crust at rifted, slowly spreading ridges such as the Mid-Atlantic Ridge. Lava flows are often thinner along the much faster spreading East Pacific Rise and, therefore, more eruptive events may be required to build the same length of crust there. This speculation, though, has not yet been confirmed by drilling.

Coarse-grained gabbros crystallized from the magma chamber complex comprise approximately two-thirds of ocean crust. They may exist as solids in the deep crust due to the hydrothermal processes acting at the top of the crust that cool the steady-state crystallizing mass to nearly 90°F (50°C) below the average temperature of magmas arriving from the mantle. This cooling mechanism will crystallize at least 50–70% of molten material at a fast-spreading ridge before it can erupt. At slowly spreading ridges, where magma is supplied more gradually, hydrothermal cooling drives crystallization in the deep crust almost to completion. At both rapidly and slowly spreading ridges, crystallization occurs by the plating out of minerals (olivine, plagioclase, and clinopyroxene) along the walls of intrusive bodies in the gabbroic mass, all within close range (a few kilometers) of the center of spreading. The crystallization forces changes in the compositions of melts and thus drives magmatic differentiation in the ocean crust. The sequence of liquid compositions produced by this process changes from olivine basalt, to titanium-rich ferrobasalt, to rarely erupted ferroandesites and dacites. The latter liquids often remain in the lower crust, crystallizing to oxide-rich gabbro and plagiogranite. Occasionally, at slowly spreading ridges, rift dynamics will allow normal faults to penetrate the bodies of crystallizing magma-chamber complexes. This produces shear zones along which melts may migrate. Thus, dredged and drilled gabbroic rocks from these environments are often lithologically complex. While the deformed rocks may be highly sheared, with gneissic or porphyroclastic textures, they are often cut by younger veins of oxide gabbro or granitic material. *See* DACITE; GABBRO; GNEISS; OLIVINE.

The fundamental melt product at mid-ocean ridges is variously termed abyssal tholeiite or mid-ocean-ridge basalt. In most places, this parental material is remarkably uniform in its fundamental petrologic and geochemical properties. Thus, it has a consistent crystallization sequence from place to place, and in most areas possesses depleted isotopic and trace-element signatures. The source of most abyssal tholeiites is a vast, globally distributed shallow mantle reservoir that previously experienced partial melting (thus losing radioactive parents to the radiogenic isotopes) at some distant time in the geological past. Currently, most tholeiites are experiencing further partially melting. There are places along spreading ridges, however, where less depleted or even enriched basaltic lavas are erupting. Most of these are in elevated, hot-spot regions such as Iceland and the Azores in the North Atlantic, Bouvet in the South Atlantic, Reunion in the Indian Ocean, and Easter Island in the Pacific. The ocean crust near these islands tends to be both shallower and thicker than average, in addition to being composed of isotopically distinctive basalts. The general processes of crustal accretion at these places, though, are approximately the same as elsewhere in the ocean crust. *See* EARTH; FAULT AND FAULT STRUCTURES; ISOSTASY; STRUCTURAL GEOLOGY; SUBDUCTION ZONES. James H. Natland; Walter D. Mooney

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Earth interior

All of the Earth beneath the land surface and the ocean bottom, including the crust, the mantle, and the core. The interior is not accessible to direct observation. Nevertheless, a rather detailed model has been constructed on the basis of measurements made at or above the surface. Measurements of gravity, the geomagnetic field, surface heat flow, and surface deformation can all be used to put constraints on the Earth model, but the most detailed information about the interior is provided by seismic measurements. To the nonspecialist, seismic methods are perhaps best known for their application to the oil exploration industry, where seismic data are used to map the subsurface structure of sedimentary basins. In the exploration of the Earth's interior, the seismic waves being analyzed are usually generated by earthquakes, and measurements are made of waves propagating through the interior of the body (body waves), waves propagating along the surface (surface waves), and standing waves bringing the whole Earth into a state of oscillation (free oscillations). Such measurements, when properly interpreted, provide information about seismic-wave velocities in the Earth. On the other hand, seismic-wave velocities can also be measured in laboratory experiments where rock samples are subjected to the high pressures and temperatures typical of conditions in the deep interior. Meteorites provide rock samples of materials that are probably abundant in the solar system. The comparison of laboratory and field measurements thus leads, by inference, to a model where the composition and temperature distribution can be specified to some extent. What emerges from these studies is a picture not only of the structure but also of the evolution and dynamics of the Earth's interior.

Earth among the planets. The Earth is a planet with a mean radius of 3959 mi (6371 km). The principal departure of the Earth from a spherical shape is an equatorial bulge caused by the centrifugal effect of the Earth's rotation. The geometrical flattening, which is defined by the ratio $(a - c)/a$, where a and c are the equatorial and polar radius, respectively, is estimated to be about 1 part in 298. Internal layers similarly exhibit a flattening, which gradually decreases toward the center of the Earth. The mean density of the Earth is 3.189 oz/in.^{3} (5.517 g/cm^3), and the moment of inertia is 0.527 times that of a body of constant density with the same size and mass as the Earth. The fact that this factor for the Earth is much smaller than unity is one of the clearest indications that the Earth has a central core of greater density than the surrounding mantle. A plausible assumption consistent with the composition of meteorites is that of a stony mantle composed mainly of minerals rich in silicate and magnesium and a core composed mainly of iron. The Earth is probably not unique in that respect. It is generally assumed that the planets and meteorites formed at the same time about 4.5 billion years ago, shortly after the origin of the Sun. From the matter that was then available, most of the lighter gaseous material was amassed by the large outermost planets, while the four smaller innermost earthlike planets and the meteorites consist mainly of heavier rocky material. The mass and moment of inertia of the earthlike planets are all consistent with the model of an iron-rich core surrounded by a more silicate-rich mantle, although the relative proportions may be different for different planets, as shown in **Fig. 1**. Venus is the planet resembling Earth most closely in size and mass, yet there are observable differences in behavior of the two planets. Of interest for comparison with the Earth's interior is the virtual absence of a magnetic field produced by Venus. *See* SOLAR SYSTEM.

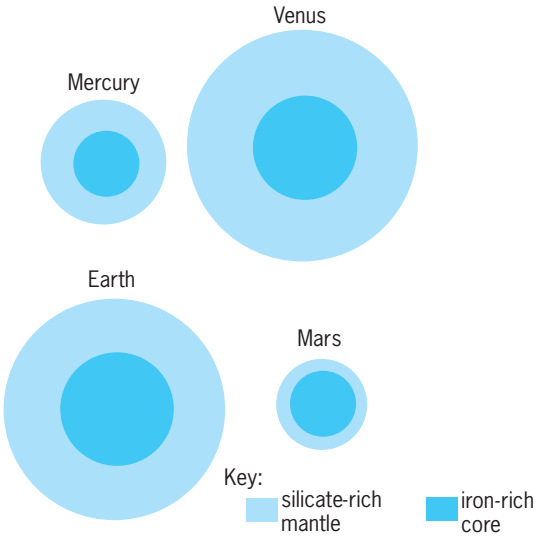

Fig. 1. Speculative models of the interiors of the terrestrial planets. All are rocky, divided into a crust (not shown here), mantle, and metallic core. (*After R. T. Merrill and M. W. McElhinny, The Earth's Magnetic Field, Academic Press, 1983*)

Earth model concept. One significant aspect of the evolution of the Earth is the differentiation of relatively heavy and light material, as evidenced in a crude way by the moment of inertia. It is then not unreasonable to envisage an idealized concept of the Earth that is spherically stratified, with the heavier material in the innermost shells. Of course, the Earth is not spherically symmetric, and this is most obvious at the surface, but a spherically symmetric model does serve the purpose of identifying the essentially layered structure of the Earth, and it serves as a reference against which to measure lateral variations of structure. Evidence for this layered structure has come first of all from seismic studies. Thus several standard Earth models have been constructed in the past. One of the more comprehensive models was constructed in 1981 by A. Dziewonski and D. Anderson, the Preliminary Reference Earth Model (PREM). The principal regions of this model are listed in **Table 1**. *See* EARTH.

Seismic structure. The seismic structure of the Earth is studied by analyzing body waves, surface waves and free oscillations, and anisotropy.

Body waves. These are generated by earthquakes or large (nuclear) explosions and are recorded by many seismic stations throughout the world. There are about 2000 stations, but any particular wave is usually recorded at only a limited subset of these. Body waves are characterized as P (primary) and S (secondary) waves. Both wave types are supported by a solid, but P waves have a higher velocity and arrive first. It is routine practice that the stations report the arrival time of the first P wave to an earthquake information center, where these data are used to locate the earthquake and determine its origin time. By subtracting the origin time from the arrival times, the travel times of waves from the earthquake to the stations are determined, and these travel times can be used to determine the seismic velocity structure of the Earth's interior. Methods to determine the three-dimensional velocity structure are collectively known as tomography, a term derived from similar applications of travel time inversion used in medicine. However, the original purpose in seismology was to determine the velocity structure of a spherically stratified Earth model. The pioneering work of H. Jeffreys and K. Bullen in 1932–1942 resulted in travel time tables for many types of P and S waves, and formed the basis for the construction of P and S velocity models of the Earth. *See* EARTHQUAKE; SEISMOLOGY.

Surface waves and free oscillations. These are also generated by earthquakes. The velocity with which surface waves propagate is a function of the frequency of the waves, that is, the surface waves are dispersive. The dispersion depends on the seismic velocity structure of the surface layers. The free oscillations of the Earth resonate only at certain discrete frequencies (the eigenfrequencies), that is, the free oscillation spectrum would be a line spectrum (in practice the lines are somewhat smeared because of dissipation of the oscillations). The eigenfrequencies depend on the velocity structure of the whole Earth. Thus surface waves and free oscillations provide an alternative data set for constructing velocity

TABLE 1. Principal regions of a standard Earth model

	Layer	Approximate depth range,* mi (km)	
(1)	Ocean layer	0–1.8	(0–3)
(2)	Upper and lower crust	1.8–15	(3–24)
(3)	Lithosphere below the crust	15–50	(24–80)
(4)	Asthenosphere (low-velocity zone)	50–140	(80–220)
(5)	Upper mantle above phase or compositional changes near 240 mi (400 km)	140–240	(220–400)
(6)	Transition region between phase or compositional changes near 240 and 416 mi (400 and 670 km)	240–416	(400–670)
(7)	Lower mantle above core–mantle boundary layer	416–1703	(670–2741)
(8)	Core–mantle boundary layer (D″)	1703–1796	(2741–2891)
(9)	Outer core	1796–3200	(2891–5150)
(10)	Inner core	3200–3959	(5150–6371)

*Depth ranges are uncertain, especially in the crust and upper mantle.

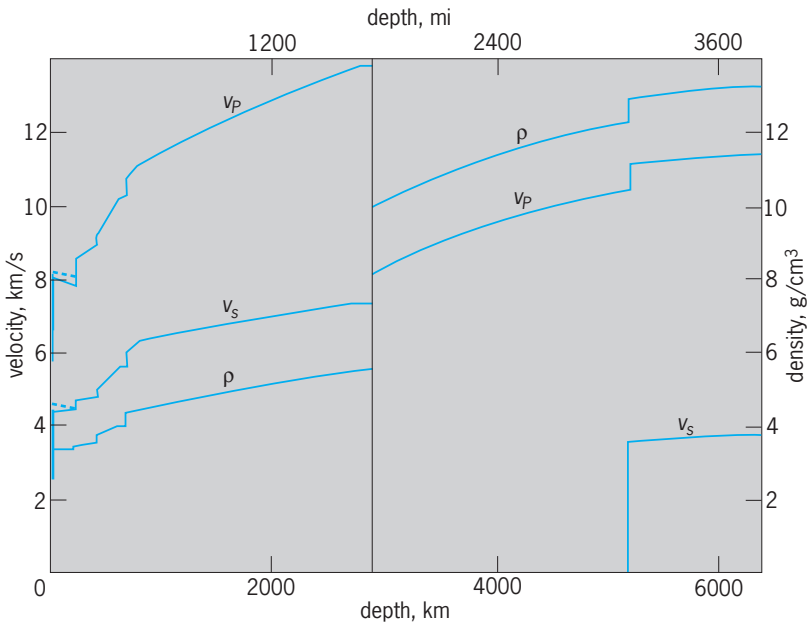

Fig. 2. Preliminary Reference Earth Model (PREM) for P- and S-wave velocities, v_p and v_s, and density ρ. In the part of the upper mantle where the model is anisotropic, velocities are given in both the vertical and horizontal direction. The broken lines represent the horizontal components of velocity. 1 km = 0.6 mi. 1 g/cm^3 = 0.5902 oz/in.^{3} (*After A. M. Dziewonski and D. L. Anderson, Phys. Earth Planet. Int., 25:297–356, 1981*)

models. After the big Chilean earthquake of 1960 provided the first observational evidence of free oscillations of the Earth, methods for inverting eigenfrequencies were developed. However, when seismic velocity models based on free oscillations became available, it was discovered that they were somewhat slower than velocity models based on body-wave travel times. It was initially thought that the travel time model was biased, since nearly all stations are located on continents. However, the effect was not sufficient to explain the difference, and the question was first resolved when the effect of anelastic damping of the waves was taken into account. PREM is based on both travel times and free-oscillation data. The model for a wave period of 1 s is shown in **Fig. 2**. *See* OSCILLATION.

The quality factor (Q) is used to quantify the damping of wave motion. The reciprocal of the quality factor (Q^{-1}) is defined as the fraction of elastic energy dissipated, per wavelength for traveling waves like body waves or surface waves, and per period for standing waves like free oscillations. Two consequences of the dissipation are a slight shift of the eigenfrequencies of standing waves and a slight change of the velocity of traveling waves. The changes depend on the period of the waves. In particular, a velocity model based on free oscillations with periods about 200 s is expected to be slower than a model based on body waves with periods about 1 s, which is precisely what the results show.

The quantity Q is also an important indicator of the physical state of the interior. The intrinsic Q at any point equals the Q of a body wave passing through. There is in fact a different Q for P and S waves; usually S waves dissipate faster than P waves, hence $Q_s < Q_p$. Q is a function of frequency such that dissipation effectively takes place within a certain frequency band, called the absorption band. The position of this band depends on temperature, pressure, and other special conditions. **Table 2** gives a rough estimate of Q for P waves at a 1-s period in different regions of the Earth. Q is strikingly low in the asthenosphere and near the top of the inner core. These are also solid regions of the Earth believed to be near the melting point. Viscosity similarly decreases with increasing temperature, and a connection between Q and viscosity has been contemplated. *See* ASTHENOSPHERE; VISCOSITY; WAVE MOTION.

Anisotropy. This refers to the directional dependence of the wave velocity. Anisotropy in the Earth's interior may be induced by flow of material containing noncubic crystals or elongated grains. Anisotropy has been observed in the lithosphere, especially under the oceans, and has been postulated for the asthenosphere. This anisotropy has been explained in terms of alignment of olivine crystals, which are thought to be the dominant material in these layers. In PREM the asthenosphere and lithosphere below the crust are anisotropic, although for practical reasons the anisotropy is of the simplest possible type such that there is no directional dependence within the horizontal plane; this is known as uniaxial anisotropy or transverse isotropy. The anisotropic regions of PREM (Fig. 2) are characterized by velocities in both the vertical and the horizontal directions. Anisotropy may occur in deeper regions of the Earth as well, but is difficult to resolve. However, to account for the differences in travel time between P waves passing through the inner core in the equatorial and polar directions, respectively, it has been proposed that the inner core is anisotropic, although the physical mechanism is not understood. *See* LITHOSPHERE; OLIVINE.

TABLE 2. Q for P waves at 1-s period

Region	Depth range, mi (km)	Q_p
Lithosphere	0–50 (3–80)	1500
Asthenosphere	50–140 (80–220)	200
Upper mantle	140–240 (220–670)	400
Lower mantle	416–1796 (670–2891)	1000
Outer core	1796–3200 (2891–5150)	10,000
Top of inner core	3200–3262 (5150–5250)	200
	3262–3387 (5250–5450)	300
	3387–3511 (5450–5650)	600
Central inner core	3511–3959 (5650–6371)	1000

Density structure. In an elastic body like the Earth, the seismic velocities v_p and v_s at any point depend on the local elastic moduli and the density. For most purposes, just two elastic moduli are needed: the incompressibility or bulk modulus, and the rigidity or shear modulus. The incompressibility measures the resistance to volume change under hydrostatic pressure, and the rigidity measures the resistance to change of shape under shear stress. Since there are just two measured quantities, v_p and v_s, the three parameters of the medium—density, bulk, and shear modulus—cannot be determined without additional assumptions and constraints. Two constraints are given by the Earth's mean density and moment of inertia. Another experimental constraint is given by laboratory measurements of the relation between v_p and density for relevant rock samples. Further assumptions are that in each of the principal layers of the Earth, where the seismic velocity profile is smooth, the composition of the material is chemically homogeneous, and phase changes are absent. The density distribution in each layer can then be computed by a method developed by L. Adams and E. Williamson, with a correction term added by F. Birch in case the temperature gradient exceeds the adiabatic value. Bullen, in the 1930s, used these principles with the assumption of an adiabatic temperature gradient in the different layers (so that there is no correction term) to construct a density structure consistent with the seismic velocity model. Later in the 1970s, this density model could be tested independently by using free-oscillation eigenfrequencies. It is necessary to distinguish between toroidal and spheroidal oscillations. Particle motion in the latter has a nonzero radial component, and the eigenfrequencies depend on the density structure. The density structure needed to satisfy the free-oscillation data is remarkably similar to Bullen's original density model. The density structure of PREM is shown together with the seismic velocities in Fig. 2. For the core and lower mantle, deviations from homogeneous and adiabatic conditions are insignificant, except for the region immediately above the core–mantle boundary.

The incompressibility, the rigidity, the gravity, and the hydrostatic pressure inside the Earth are obtained as by-products of the density calculations. The incompressibility increases steadily with depth throughout the mantle and core, and its changes across the core–mantle boundary are remarkably small. The rigidity also increases steadily throughout the mantle, but the fluid outer core has zero rigidity. In the inner core the rigidity is again nonzero but is probably less than in the lower mantle. This gives an abnormally low ratio of shear to bulk modulus for the inner core. The variation of gravity and pressure inside the Earth is illustrated in **Fig. 3**. *See* ELASTICITY.

Composition and state. Estimated abundances of rocky, refractory materials in meteorites and the Sun give a rough idea of what the Earth is composed of, although the relative proportions of metals and minerals in various portions of the Earth are not necessarily identical. To find out where and in which proportions the different materials reside in the Earth, high-pressure and high-temperature laboratory measurements are compared to the seismic and density structure. The Earth comprises a crust, a mantle, and a core, so that there is a compositional differentiation into at least three regions. Each of these regions is differentiated again, both vertically and, at least for the crust and upper part of the mantle, laterally. *See* ELEMENTS, GEOCHEMICAL DISTRIBUTION OF; HIGH-PRESSURE PHYSICS.

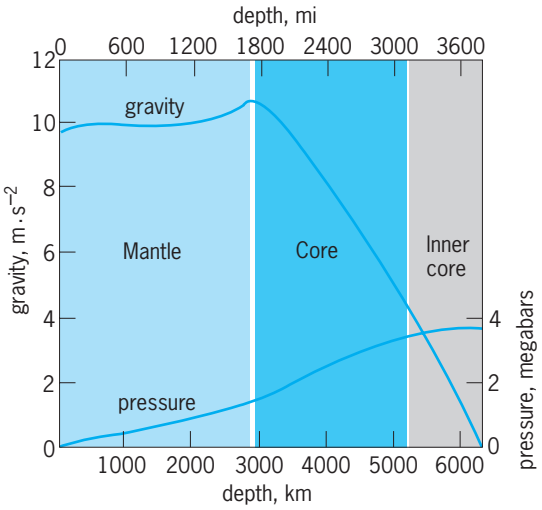

Fig. 3. Variations of gravity and pressure inside the Earth. 1 m = 3.3 ft. 1 megabar = 100 GPa. (*After P. Melchior, The Physics of the Earth's Core, Pergamon Press, 1986*)

Crust. The Earth's crust is on average about 12 mi (20 km) thick but is thinner under oceans and thicker under continents. Thickness under some high mountain ranges may be up to 36 mi (60 km; the so-called mountain roots). The boundary between the crust and the underlying mantle is called the Mohorovičić discontinuity or, more usually, the Moho. The continental and oceanic portions of the crust have different compositions. Continental crust can be divided into a lighter, granitic upper crust and a lower crust that is more basaltic, like the oceanic crust. Despite the striking heterogeneity of the crust, some generalizations can be made. Both the crust and the

TABLE 3. Chondritic and solar estimates of terrestrial composition versus upper-mantle composition (percent by mass)*

Component	Chondritic	Solar	Lherzolite†
Mantle			
MgO	37.7	32.7	42.2
SiO_2	52.5	45.0	44.2
Al_2O_3	3.8	3.2	2.1
CaO	3.0	3.4	1.9
FeO	3.3	15.7	8.3
MgO/SiO_2 ratio	0.718	0.727	0.955
Core			
Fe_2O	32.2	32.0	

*After D. L. Anderson, Where on Earth is the crust?, *Phys. Today*, 42:38–46, 1989.
† Composed mainly of olivine. Believed to be the major rock type in shallow upper mantle.

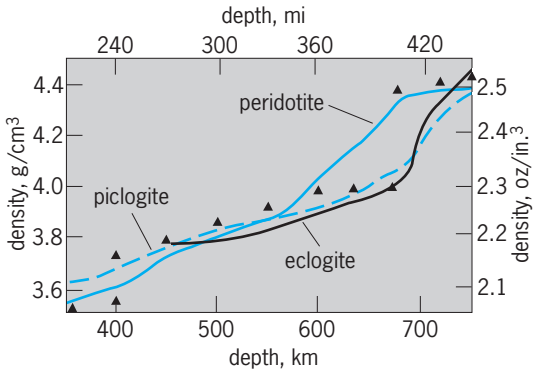

Fig. 4. Density versus depth for several mantle constituents along the 1400°C (2552°F) adiabat. Data points show the density variation of PREM. All three constituents shown here contain magnesium-rich silicates. (*After D. L. Anderson, Where on Earth is the crust?, Phys. Today, 42:38–46, 1989*)

mantle are rich in silicates, but the mantle has much more magnesium, and the crust has proportionally more of the relatively light elements such as calcium, aluminum, sodium, and potassium in the form of oxides. It has been noted that the relative abundance of these light elements in the stony meteorites leads to an expectation that much more of these elements should be on Earth than is contained in the crust. The missing part may be contained in former crust that has been subducted in the process of plate tectonics and is now on its way to being recycled throughout the mantle. Only oceanic crust is being subducted. *See* EARTH, CONVECTION IN; EARTH CRUST; METEORITE; MOHO (MOHOROVIČIĆ DISCONTINUITY); PLATE TECTONICS.

Mantle. This is the region between the crust and the core. It is common to make a distinction between the upper mantle and lower mantle, which are separated by the so-called 670-km discontinuity at a depth of 416 mi.

The upper mantle includes the lithosphere below the crust, the asthenosphere, and the region of phase transformations (the transition zone). It contains also the subducting slabs of oceanic lithosphere. The difference between continental and oceanic crust extends to at least the bottom of the lithosphere and probably to the bottom of the asthenosphere. It is believed that the stiff lithosphere and the soft asthenosphere differ mainly in mechanical strength, not in composition. The shallow part of the upper mantle is likely to be relatively depleted in the light elements typical of the crust. A comparison with estimates for stony meteorites (chondrites) and for the Sun (**Table 3**) serves to illustrate the depletion. On the other hand, the light elements do not seem to be underrepresented in estimates of the composition of the deeper parts of the upper mantle. It has been suggested that this region contains more eclogite, the high-pressure form of basalt, which is subducted with the oceanic lithosphere. It is the densest common upper-mantle mineral. Olivine is probably the most abundant upper-mantle mineral. At pressures prevailing near 240 mi (400 km) depth, it transforms to a denser structure (spinel), with a density increment of about 10%. Phase transformations of other common upper-mantle minerals are also expected at the depths spanned by the transition zone (**Fig. 4**). *See* BASALT; ECLOGITE; SPINEL.

The lower mantle, the mantle below 416 mi (670 km), appears to have a homogeneous composition down to the boundary layer above the core. Experimental results seem to indicate that mineralogy

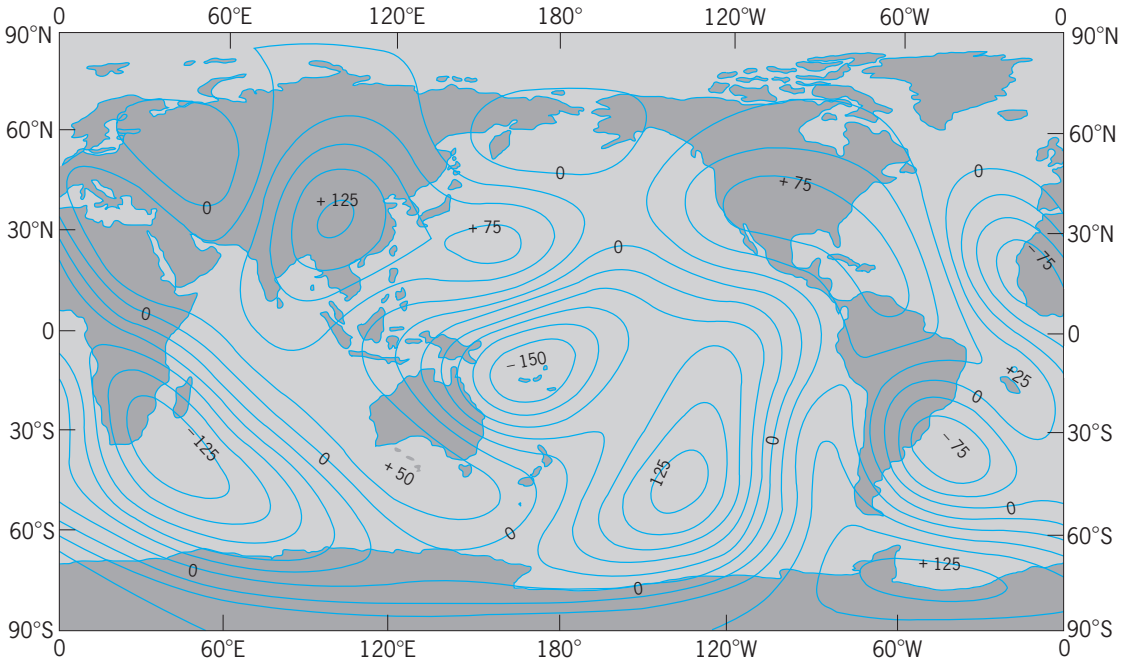

Fig. 5. P-wave velocity anomalies above the core–mantle boundary in terms of spherical harmonics up to degree and order 6, obtained by inverting P-wave travel time data. Positive perturbations indicate faster-than-average material. Contour interval = 25 m/s; 1 m = 3.3 ft. (*After A. M. Dziewonski, Determination of lateral heterogenicity in P velocity up to degree and order 6, J. Geophys. Res., 89:5929–5952, 1984*)

is dominated by ferromagnesium oxides (wüstite) and silicates (perovskite). The high-pressure perovskite phase of ferromagnesium silicate is thought to be the most abundant mineral of the Earth. *See* PEROVSKITE.

670-km discontinuity. This represents the transition from upper to lower mantle. There seems to be no compelling geochemical evidence against the view that the presumed dominant lower-mantle minerals are the high-pressure products of upper-mantle spinel, and there would be no chemical change across the boundary. On the other hand, such a change is not precluded by the evidence either, and an increased iron oxide content for the lower mantle is sometimes proposed. In any case, the transition represents a density increase of perhaps 10% and a viscosity increase by perhaps a factor of 10. It is not clear that subducted lithospheric slabs can penetrate this barrier. In fact, the seismic evidence suggests that some slabs do penetrate but others do not. One speculative model consistent with these results is that relatively young, thin oceanic plates are trapped at the boundary, but the upper, cool differentiated parts of thick plates may break through. It would imply a so-called leaky barrier to mantle flow.

Core–mantle boundary layer. This is often identified with the layer labeled D″ in Bullen's original seismic model. The seismic heterogeneity in this layer has a wide spectrum of scale lengths. **Figure 5** is a smoothed picture of large-scale lateral variations of P-wave velocity. Significant small-scale features are also inferred. From temperature considerations it is clear that the seismic anomalous layer is also a thermal boundary layer, and it has been proposed that this layer becomes unstable and forms convection plumes. The very big contrast in composition and material properties across the core–mantle boundary (**Table 4**) also creates favorable conditions for chemical heterogeneity above the boundary, although the observational evidence is unclear. The argument is that patches of material intrinsically denser than the bulk of the mantle will be trapped at the boundary, in much the same way as the relatively light continents are considered to be left over on the surface of the Earth. The analogy explains the wording "continents on the core–mantle boundary." However, unlike the surface of the Earth, the boundary itself is relatively smooth, with undulations of the order of a kilometer or less. This may be a consequence of low viscosity of the material at the boundary. Ironically, although it might seem that seismology would put strong constraints on topography of the boundary, the initial maps showed very large undulations. With the results of a more recent seismological study in agreement with a relatively smooth boundary but requiring variations in the boundary layer, the paradox appears to be removed. A speculative diagram synthesizing the various pieces of information is shown in **Fig. 6**.

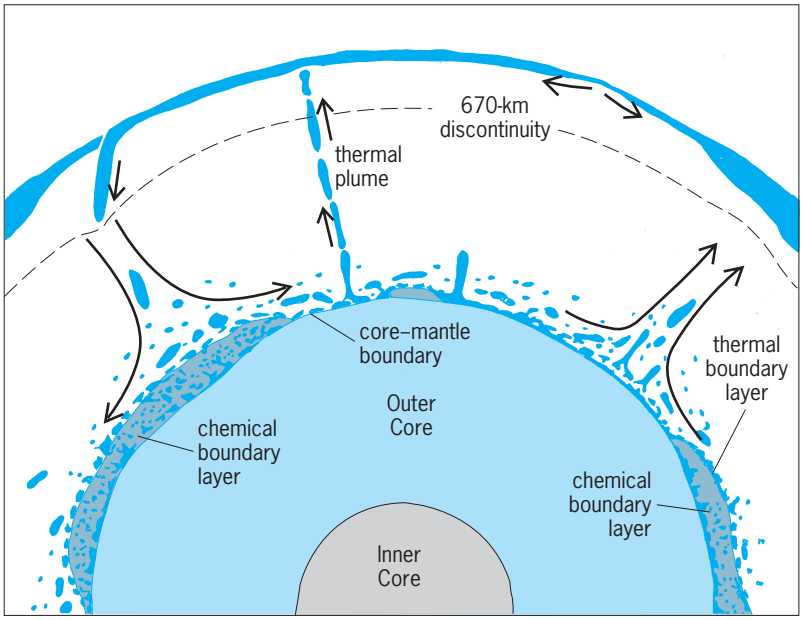

Fig. 6. Speculative, schematic model of the core–mantle boundary layer. A heterogeneous chemical boundary layer is embedded in a thermal boundary layer. Large-scale mantle circulation transports chemical heterogeneity to the base of the mantle as well as returning some by entrainment. Thermal plumes caused by boundary-layer instabilities ascend from the transition zone and disrupt the chemical boundary layer. Although the core–mantle boundary itself is relatively smooth, some topography is induced by both the large- and small-scale dynamic systems. 670 km = 400 mi. (*After T. Lay, Structure of the core–mantle transition zone: A chemical and thermal boundary layer, Eos, 70:49–59, 1989*)

Core. This is the central region of the Earth. Its seismic and density structure are well matched by a solid inner core of iron or nickel iron and a liquid outer core of iron mixed with about 10% lighter material. Candidate light elements are sulfur and oxygen, with some preference for the former. The outer core can be taken as homogeneous for all practical purposes. Seismic Q suggests that the inner core boundary is at the melting point. This is not the melting point of

TABLE 4. Contrast in material properties across the core–mantle transition zone*

Composition	Iron + 10% (O, S, Si, C)†	Magnesium-iron oxides and silicates
State	Molten alloy	Subsolidus
Temperature	3800–4700 K (6380–8000°F)	2600–3100 K (4680–5120°F)
Flow velocity	$0(10^{1})$ km/year	$0(10^{-4})$ km/year
Viscosity	$0(10^{-2}–10^{3})$ Pa · s	$10^{22}–10^{23}$ Pa · s
Density	9.90×10^{3} kg/m^3	5.57×10^{3} kg/m^3
Rigidity	0(0?) Pa	2.911×10^{11} Pa
Electrical conductivity	$10^{5}–10^{6}$ S/m	$10^{-2}–10^{2}$ S/m
Incompressibility	6.35×10^{11} Pa	6.85×10^{11} Pa

*After T. Lay, Structure of the core–mantle transition zone: A chemical and thermal boundary layer, *Eos*, 70:49–59, 1989.
†O = oxygen; S = sulfur; Si = silicon; C = carbon.

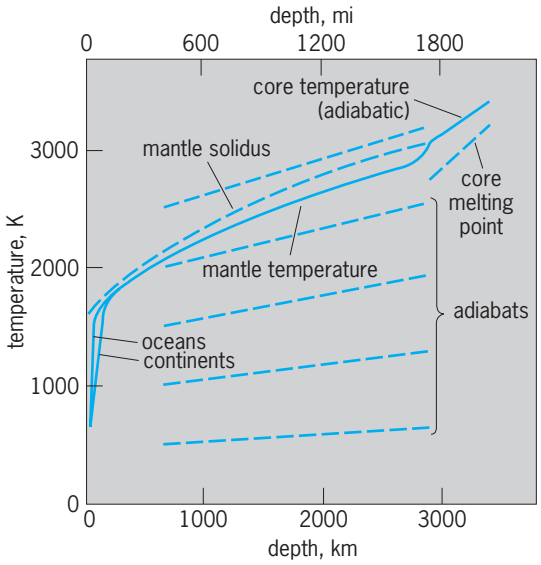

Fig. 7. Schematic temperature profile of the mantle. Subsequent estimates of core temperature are higher than suggested in the figure; the temperature jump across the thermal boundary layer at the base of the mantle will then be correspondingly higher. °F = (K × 1.8) − 459.67. (*After F. Stacey, Physics of the Earth, 2d ed., Wiley, 1977*)

pure iron; the presence of impurities in the liquid depresses the melting point considerably.

Core–mantle coupling. This is believed to be responsible for decade-long variations in the length of the day of up to about 5 milliseconds. The idea is that fluid flow in the outermost regions of the core is coupled to the mantle, so that the rotation rate of the mantle can fluctuate in response to time variations of the flow in the core. This view is supported by an observed correlation between the decade variations in the length of the day and certain variations of the geomagnetic field (the latter being associated with core flow). The coupling may be electromagnetic or topographic. It has been suggested that topographic coupling can be quite effective, provided that there is indeed topography on the core–mantle boundary (of the order of a few hundred meters).

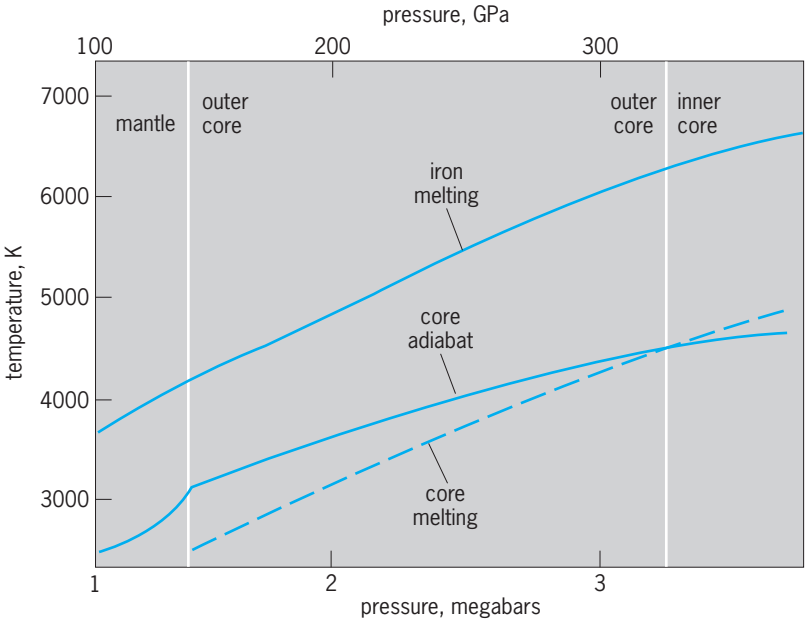

Fig. 8. Schematic temperature profile of the core. The melting-point curve for the core is substantially below the melting point of pure iron. All temperatures may have to be raised to correspond with further estimates of the melting point of iron. °F = (K × 1.8) − 459.67. (*After D. J. Stevenson, Models of the Earth's core, Science, 214:611–619, 1981*)

Temperature. Mantle temperature estimates are constrained by the following: (1) A melting-point curve for the mantle is obtained from laboratory measurements of the melting point of presumed mantle minerals under pressure. The actual temperature should not be too far below the melting point in order that material can flow. (2) Convection is supposed to be the dominant mode of heat transport below the lithosphere. In a slowly convecting system the temperature gradient is probably slightly superadiabatic, which is a gradient larger than that prevailing under adiabatic conditions. (3) The temperature gradients in the crust can be measured. Since conduction is the mode of heat transport through the lithosphere, temperatures can be estimated by extrapolating the surface temperature downward, until the melting point curve for the mantle is approached. This point is often taken as the bottom of the lithosphere, and temperatures in the mantle are estimated by extrapolating downward from this point; a schematic temperature profile is shown in **Fig. 7**. The thermal gradients on continents and under oceans differ substantially, and continental lithosphere is thicker than oceanic. At the base of the mantle a thermal boundary layer with a strongly superadiabatic gradient exists. Estimates of core temperature suggest that the temperature jump across this layer is larger than is indicated in Fig. 7. The figure does not suggest a thermal boundary layer between the upper and lower mantle (near 416 mi or 670 km depth); but this is, in fact, an unresolved problem.

Core temperature presumably follows a nearly adiabatic gradient, at least in the convecting outer core. A calibration point for the temperature profile is provided by laboratory measurements of the melting point of iron under high pressure and by assuming the inner core boundary is at the melting point. This point is about 1000 K (1800°F) below the melting point of pure iron under the same pressure, since the liquid iron (in the outer core) is not pure but has about 10% light material mixed in. A temperature profile for the core is shown in **Fig. 8**. More recent estimates of the melting point of iron under high pressure are higher than suggested in Fig. 8, and so the actual temperature profile will be correspondingly higher.

Heat sources are the radioactive elements, especially in the continental crust and the upper mantle, and the slowly cooling Earth itself. The total surface heat flux is about 3×10^{13} W, and cooling may contribute about a quarter of this value. A flux of about 3×10^{12} W through the core–mantle boundary may be due to cooling of the core and the consequent crystallization of the inner core, although the possibility of heating due to decay of radioactive potassium (^{40}K) cannot be neglected. The crystallization

process itself is a good example of ongoing differentiation of the Earth. *See* EARTH, HEAT FLOW IN; RADIOACTIVITY.

Geomagnetic dynamo. The Earth's magnetic field has been studied for centuries, but it is only rather recently that an understanding of the generation of the field has developed. In 1946 W. Elsasser proposed a model of the magnetic field generated by inductive motions in a liquid core. Evidence for a liquid core came 20 years earlier, when H. Jeffreys noted the absence of seismic S waves passing through the core. A geomagnetic dynamo theory, based on the work of W. Elsasser and later on that of E. Bullard, is probably the only theory capable of explaining the important features of the magnetic field such as reversals, secular variation, and the near alignment of the dipole and rotation axes. However, the mathematical problems are formidable, and as yet no satisfactory dynamo model exists that reproduces secular variation and reversals. An intriguing hypothesis is that these features are governed by conditions in the core–mantle boundary layer. *See* GEOMAGNETISM.

Reversals. This term refers to polarity transitions of the dipole part of the geomagnetic field. The field on the Earth's surface is about 90% dipole. Paleomagnetic results have demonstrated that the polarity of the field has reversed many times in the past, but the frequency of reversals appears to be nonstationary. One hundred million years ago the magnetic field was in the middle of a 30-million-year period of normal polarity (the Cretaceous quiet interval). This is in sharp contrast to the behavior both 150 million years ago and recently, when the mean polarity interval is of the order of several hundred thousand years. Although the connection is unclear, such long-term changes are currently believed to be related to variations in the thickness of the core–mantle boundary layer; thermal boundary-layer studies indicate that such variations can occur. It is speculated that when the layer is thick, the cooling rate and power supply are small and the field is in a quiescent state, with reversals infrequent or even absent. Conversely, when the layer is thin, the cooling rate and power supply are large and the field is in an agitated state with frequent reversals, which is the state occurring at present. *See* PALEOMAGNETISM.

Secular variation. Included in the secular variation of the magnetic field, among other factors, is the westward drift of the field and the decrease of the dipole field strength during the past few hundred years; if the decrease continues at the present rate, a reversal will occur in roughly a thousand years. The secular variation of the magnetic field at the Earth's surface has been reasonably documented during the past few hundred years. J. Bloxham and D. Gubbins have made an estimate of the field at the core–mantle boundary during this time period that implies inverting the observed data of the surface magnetic field. In **Fig. 9** the field is shown for the years 1715 and 1980. It is striking that some prominent features seem to be stationary, and the speculation is that these features are controlled by lateral temperature variations in the core–mantle boundary layer.

Energy source. The source of the energy for the geodynamo can be specified on the basis of an energy balance argument. It is estimated that the power needed for the dynamo process itself is about 4×10^{11} W. It is also estimated that in the process a heat flux equaling about 2.7×10^{12} W flows through the core–mantle boundary by conduction. An obvious heat source is radioactive decay; this requires a certain amount of radioactive potassium in the outer core. The problem with this process is the very low efficiency; only about 5.5% of the heat is becoming available to power the dynamo.

Another mechanism not suffering from an efficiency problem has been proposed. When the core cools, solid iron (or nickel-iron) freezes out on the

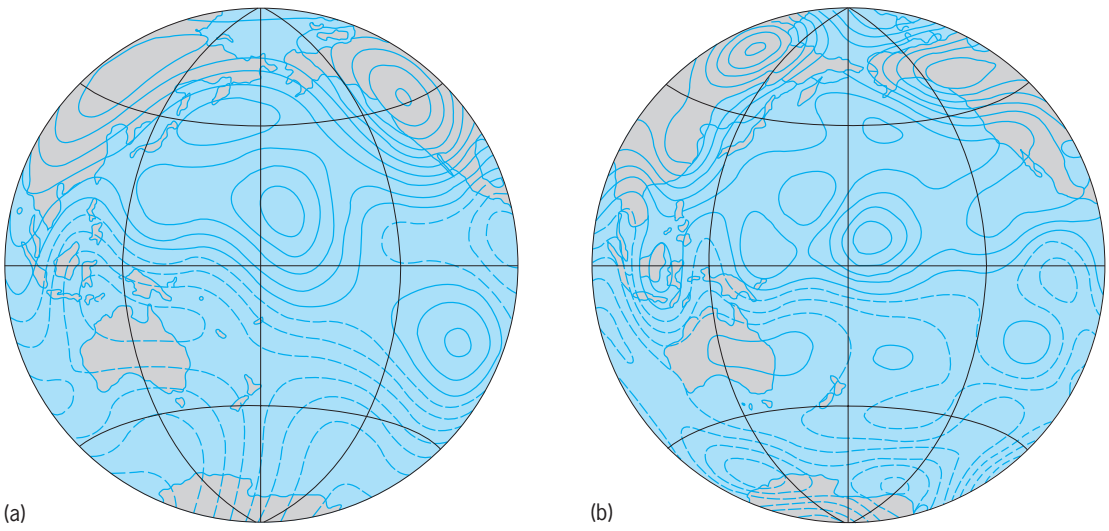

Fig. 9. Contour plots of the radial component of the magnetic field at the core–mantle boundary (*a*) for the year 1715 and (*b*) for 1980. The contour interval is 100 microteslas; solid-line contours represent flux into the core, and broken-line contours represent flux out of the core. (*After J. Bloxham and D. Gubbins, The secular variation of the Earth's core, Nature, 317:777–781, 1985*)

boundary of the inner core, leaving a lighter constituent behind in the outer core. The rising of the lighter material through the liquid core leads to so-called compositional convection, and its efficiency is very high. The additional latent heat of crystallization is less important in that respect.

Assuming the inner core has grown to its present size in about 3 billion years, the available power may exceed 4×10^{11} W if the density jump across the inner core boundary exceeds 0.302 oz/in.^{3} (0.5 g/cm^3). The density difference is thus a crucial quantity for the geodynamo. In principle, seismology can provide estimates of the density. In practice, the density jump across the inner core boundary is not well resolved.

Durk J. Doornbos

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Earth resource patterns

The physical character and distribution of natural resources at the face of the Earth. No section of the Earth is exactly like any other in its resource endowment. Combinations of land and ocean, latitudinal differences in insolation, variations in receipt of precipitation, patterns of geology, and deformation of the Earth's crust all converge to create different resources in various regions of the world. Nevertheless, there is some repetition in these natural features and forces of nature across the Earth. These similarities that exist from place to place distinguish regional patterns in the availability of resources on a global scale.

Delineation of the Earth's resource patterns begins with differentiation between continental and marine resources. Although the resources of the oceans and seas have been used by people since earliest times, the more than 6,000,000,000 on the Earth today are primarily dependent upon the resources of the land for their existence.

Five principal resources associated with the land are soils, forests, grasslands, fresh-water resources, and minerals. Although other resources such as native animal life may be of local importance, and although the very concept of "resources" has been extended to include such complexes as recreation resources, these five land resources remain of fundamental importance for the material support of human life.

Natural resources include any substance or material, produced by natural forces within the Earth system, that is used or valued by humans. Some of these resources more quickly regenerate, and are naturally

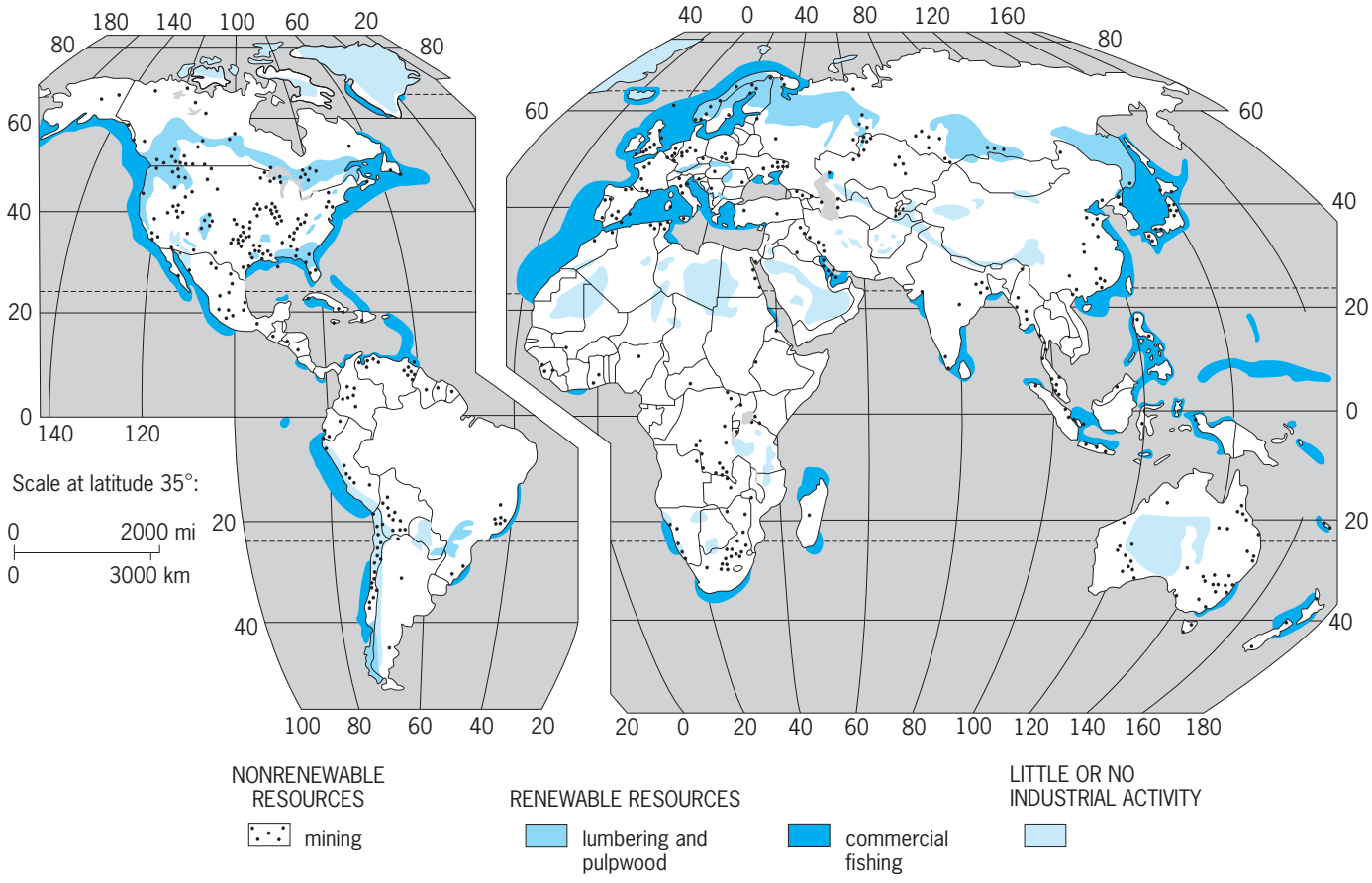

Fig. 1. Regions of selected primary and secondary industries.

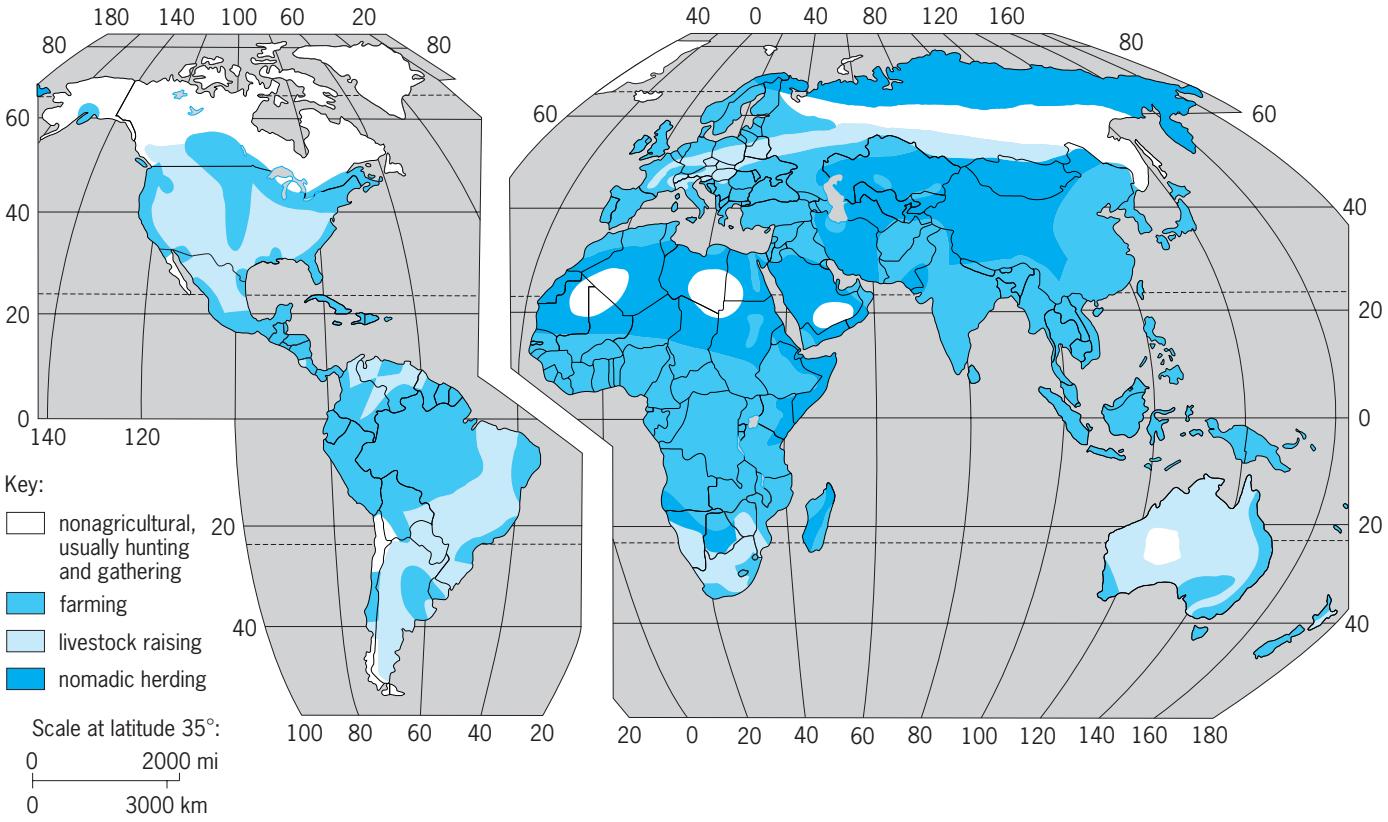

Fig. 2. Major agricultural regions.

replaced after being extracted. These resources—forests, marine life, soil, fresh water, agricultural crops—are termed renewable resources (**Figs. 1** and **2**) and may be used indefinitely if rate of use is balanced with rate of regrowth or replenishment. Other resources cannot be regenerated so quickly and are steadily used up by humans. These nonrenewable resources include oil, natural gas, coal, and metals.

Two types of natural forces have converged to create the natural materials that are used and valued by humans. One set of forces consists of the basic climatic controls, including latitude, distribution of land and water, the wind and pressure system of the rotating Earth, the major landforms of the continents, and the elevation of the land surface above sea level. A second, independent set of forces consists of the tectonic and rock-forming processes which have operated over the Earth. The climatic controls account for variations in regional climates over the continents, and these climates in turn help to shape the the unique character of forests, grasslands, and fresh-water resources, as well as some of the fundamental attributes of soils and the agriculture they support. The second set of forces may be regarded as even more fundamental since the movements of the Earth's plates and their associated continents in conjunction with the Earth's plate tectonics account not only for the position on the Earth and hence the latitudinal location of each continent, but also for the global distribution of land and water and continental landforms, all with consequences for regional climatic patterns. Moreover, rock composition and surface configuration also influence the development of forests, grasslands, water resources, and soils which alter the patterns within the climatic regions, and surface and subsurface geology are fundamental to an understanding of the global patterns of minerals on and beneath the Earth's surface.

Climatic types. Because many resources reflect the climatic regime under which they form, it is important to understand distinctions between climates, where these various climates are found, and how climates influence the natural development of resources. World climates may be broadly classified in four types, which vary according to combinations of temperature and precipitation, and vary seasonally in both phenomena due to latitude and proximity to oceans or mountains (**Fig. 3**). Distinction is made between the so-called humid climates and the water-deficient climates, with subtypes as follows:

Humid microthermal
1. Polar and ice cap
2. Tundra
3. Taiga
4. Puna

Humid mesothermal
5. Upper midlatitude
6. Humid subtropical

Humid macrothermal
7. Wet-and-dry
8. Rainforest

Water-deficient
9. Desert
10. Semiarid
11. Mediterranean

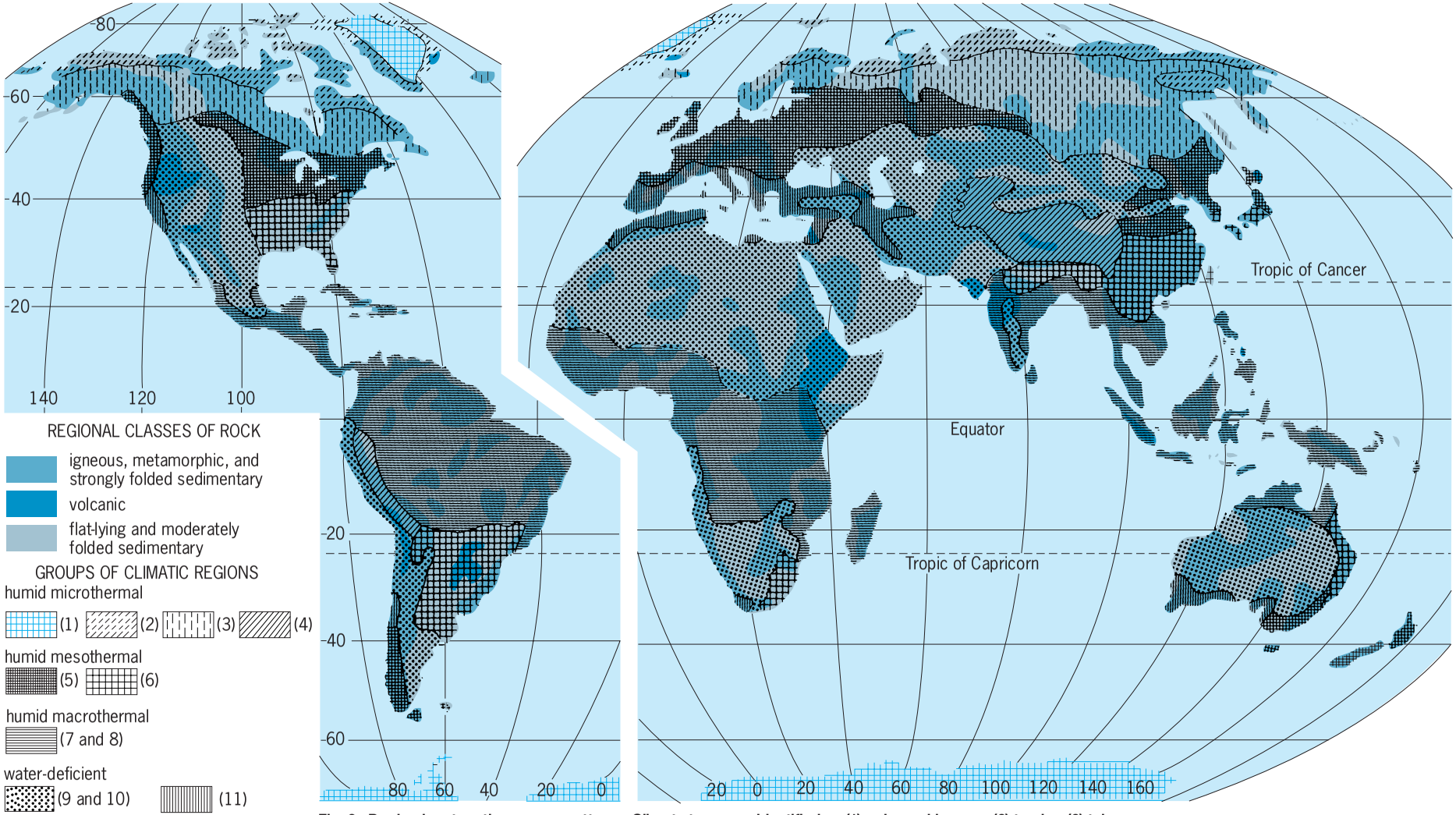

Fig. 3. Predominant earth resource patterns. Climate types are identified as (1) polar and ice cap, (2) tundra, (3) taiga, (4) puna, (5) upper midlatitude, (6) humid subtropical, (7 and 8) wet-and-dry and rainforest, (9 and 10) desert and semiarid, (11) mediterranean. The map is a flat polar quartic equal-area projection.

Humid microthermal regions. These areas of predominantly low temperature are so unfavorable to soil formation and use in agriculture that under present techniques their population-carrying capacity is low even in those regions with the warmest summers.

1. In polar and ice cap areas where soils and vegetative growth are essentially absent, available resources necessarily are dominantly marine and land animal life, on which the sparse native human settlement is almost wholly dependent. Despite the enormous size of the Antarctic, settlements are exclusively in the Arctic, except for special government-supported Antarctic stations.

2. Tundra, except for minor alpine locations, is entirely within the Northern Hemisphere. The principal renewable resources are lichens and the native animal life, such as reindeer and caribous, which can use these as food. Parts of the tundra may be considered a grazing land, as managed herds of reindeer are pastured nomadically. The natural resource significance of tundra lands for the larger world may be greatly enhanced locally where minerals are being extracted, such as the oil field adjacent to Prudhoe Bay on Alaska's North Slope. *See* TUNDRA.

3. Millions of acres of boreal coniferous forest, or taiga, located in a broad curving zone from Scandinavia across the northern, European part of Russia, and Siberia, and, east of the Bering Sea, across much of Alaska and northern Canada, has given its name to this climatic belt. Varieties of spruce, fir, and larch, which are of particular significance to the pulping industries, constitute the most valuable known renewable resource of the humid microthermal regions. The zone of taiga climate is of scant importance for agriculture; soils are predominantly thin, stony, and infertile, and development for agriculture is further discouraged by the possible incidence of frost in every month of the year and by the short growing season of 80 days or less. As in the case of the areas of tundra climate, the resource significance of particular localities within the taiga zone is locally enhanced by deposits of minerals; globally significant deposits of gold, silver, iron, lead, zinc, uranium, copper, and other metals are found in the taiga lands of Canada and Russia. Fresh-water resources are extensive, and some hydroelectric power is developed. *See* TAIGA.

4. The puna type of climate is found at much lower latitudes but at elevations generally 10,000 ft (3048 m) or more above sea level, and is characteristically cold. Certain plateaus, particularly the high plateau of Tibet and the intermontane Andean plateau (Altiplano) of southeastern Peru and western Bolivia, belong to this group. Low temperatures preclude tree growth, and the principal renewable resource is low-productivity grazing land. Thin, stony soils are limited in production to hardy small grains and root crops. The resource significance of the South American puna is greatly enhanced by metal deposits, particularly those of copper, tin, lead, and zinc. *See* PUNA.

Humid mesothermal regions. Considered in the light of present-day technology, the heart of the world's renewable natural resource base is in the humid mesothermal regions with their generally adequate precipitation and intermediate temperatures. These middle latitude regions of Earth contain a large share of the world's most productive soils which support both crop and pasture lands. Some productive coniferous and broad-leaf forests occur in these climatic regions. Concentrated surface and subsurface fresh-water supplies are relatively abundant, but owing to the high population densities and major urban and industrial complexes in these regions, the water resources are often intensively developed to the point where the quantity and quality of resources is diminished.

5. Upper midlatitude climate contains most of the lands adapted to the raising of wheat, barley, rye, and oats. In addition, certain areas within this climatic type, and particularly the North American corn belt and the Middle Danube Basin and the North Italian Plain, also produce maize. Further crops, including soybeans, are also important, and extensive forage cropping supports dairying and meat animal raising in east-central North America, western and central Europe, and the nonarid part of southern European Russia. The lands which support the agricultural types of this climatic region possess the most extensive areas of superior soils on Earth. They are relatively deep and are often moderately to highly endowed with plant nutrients and humus. Some of the best soils have developed from glacial drift; others have benefited from the deposition of fine-grained windblown loess deposits during the glacial period. Extensive preagricultural grasslands also contributed high humus content to some of the soils. Finally, the extensive plains under which these soils are deployed, and the generally favorable growing season conditions of temperature and rainfall, also underlie the present high productivity of the agriculture supported by the soils in this climatic type. It is within the upper midlatitude climate that the region of greatest surplus food production (the North American interior) exists.

6. Humid subtropical areas have an ample water supply and relatively long growing season (200 days or longer), making the best of these lands potentially very productive for crops or for forest products. Soils other than in floodplains tend to be less fertile than those of upper midlatitudes, however, owing to the effects of relatively high rainfall and temperatures on the removal of plant nutrients, and generally require fertilization for sustained cultivation over long periods. Where cropping is on alluvial soils of floodplains large and small, as in central and southern China and in southern Japan, and in the Ganges Plain of India, soils are more easily maintained at high levels of productivity. High rural population densities generally preclude the generation of large exportable food surpluses from regions with the best soils. Although the favorable combination of temperature and moisture can result in high timber growth, as on the best-managed tracts of forest in the American South, the potential is not realized in other areas of the Earth having this type of climate, owing to

previous large-scale deforestation. Water resources are generally abundant; in part of southern and eastern Asia, fresh-water resources underlie the most extensive humid land irrigation in the world—namely, for paddy rice.

Humid macrothermal regions. These low latitude, predominantly winterless regions of warm to hot temperatures are divided according to the regime of rainfall:

7. Wet-and-dry, with a pronounced dry season, characterizes the tropical savanna regions of the Earth. These tall grasslands, mixed with scattered tropical trees, offer some of the most extensive grazing lands in Africa and northern Australia. Some of the savanna region, the original home of the big game herd animals in the African plains, is now used for commercial livestock and mixed farming.

8. Rainforest, with year-round growing seasons but over extensive areas their soils are lateritic with a high iron content, and harden irreparably under use for cropping. Problems induced by fungal growth, bacterial disease, and insect abundance also handicap the use of the soil resources. Within the great alluvial valleys, flooding may also be disruptive. Owing to these adverse factors, shifting cultivation is common, and much land is not in production at any particular time. Within the rainforest regions, extensive and rapid-growing forests occur, including the largest such forest area on the Earth in the Amazon Basin, but are mostly unexploited commercially in the twentieth-century economy. Extensive clearing of the Amazon forest for agricultural development appears likely during the last two decades of the twentieth century. Sites of enormous potential hydroelectric generation are largely undeveloped. Extensive iron and aluminum deposits have become concentrated within the rainforest soils after centuries of heavy rainfall has leached away most other soil substances. Many of these deposits have been exploited by domestic and multinational mining corporations. *See* RAINFOREST.

Water-deficient regions. Receipts of moisture are scant or lacking during much of the year. Except where exotic water supplies are available for irrigation, soils inherently are less productive than those of any humid region with comparable land surface. There are three major subdivisions.

9. Deserts differ strikingly in their form and in temperature conditions, but everywhere present meager resources for agriculture. Where water is available, desert oases blossom, but vast areas contain only scrub growth, ephemerals, or virtually no vegetation. Livestock-carrying capacity of desert scrub is meager, though all but the most extreme deserts have traditionally been occupied by groups of nomadic pastoralists who typically herd goats, sheep, or camels. Sparsity of vegetation has made mineral prospecting and exploration somewhat easier than in vegetation-covered humid areas, and in this century desert occupancy often started with mineral discoveries. *See* DESERT.

10. The semiarid regions are basically grasslands, which have grazing as their characteristic resource use. Because of cyclic rainfall variability, people have converted the inherently fertile soils, as in China and the United States, to cereal growing during periods of higher rainfall. Rainfall fluctuations, however, make the soils unstable under cultivation. For this reason, these regions have suffered consistently from unsuited cultivation and overgrazing. Semiarid lands are responsive to and most productive under irrigation, but neither surface-water nor ground-water resources are adequate to irrigate more than a fraction of the area. *See* GRASSLAND ECOSYSTEM.

11. Regions of mediterranean climate, because of their winter rainfall, generally are classed as humid lands. However, the greater part of the gowing season is water-deficient, and the most productive agricultural lands depend on irrigation. Soils are the major resource since water deficiency is pronounced enough to discourage forest productivity. Major mineral deposits may complement agricultural lands in a few areas. *See* CLIMATOLOGY.

Rock composition and surface configuration. Imposed on the basic resource pattern induced by climatic differences are variations in rock composition and surface configuration which cause intraregional differences within the patterns already described. Although not exactly the same in their effects, the variations caused by these two geographical elements are often concomitant and may be treated together as follows:

1. Rock composition and structure
   a. Flat-lying and moderately folded sedimentary rocks
   b. Igneous, metamorphic, strongly folded sedimentary rocks
   c. Volcanic rocks
2. Surface configuration
   a. Floodplains and other flat or gently sloped surfaces
   b. Mountains and maturely dissected hill lands, plateau faces, of faces of cuestas

These elements of crustal variation produce the six following geographical differences in resource endowment:

First, all major agricultural lands are on flat-lying or moderately folded sediments and have gentle slopes, well exemplified in such alluvial valleys as the Mississippi, Nile, Huang, and Ganges-Brahmaputra, and such other outstanding agricultural areas as the North American corn belt or the Paris Basin.

Second, productive secondary agricultural lands, particularly in the tropics, are located on volcanic areas where soils have been formed through weathering or wind action. Examples include the Deccan Plateau in west-central peninsular India, and volcanic soils fringing much of the central mountain backbone of Java.

Third, agricultural lands are extremely limited on igneous rock areas, no matter what the surface configuration, in regions north of 40°N latitude, as illustrated in the Laurentian Shield area of Canada and

the Fenno-Scandian Shield in northern Europe. In the humid tropical and subtropical climates, however, where weathering has proceeded long enough to produce a substantial soil mantle as on the Piedmont of the southeastern United States, underlying igneous rocks do not have the same negative effect on agricultural development.

Fourth, forest lands are not limited in their extent (although limited in productivity) by either crustal rock composition or surface configuration.

Fifth, mountains are important catchment areas and sources of fresh water and the services which may be derived from water. Most hydroelectric generation, or generation potential, is associated with mountains. In arid and semiarid regions, mountains are sources of water for irrigation, domestic and industrial supply, wood products, and warm-season grazing lands.

Sixth, mineral resources have definite patterns which are associated with rock structure and composition. Major deposits of coal and lignite, petroleum, and natural gas are, with few exceptions, found in flat-lying or gently folded and faulted sedimentary rocks, as in Texan oil fields and the coal fields of the Allegheny Plateau. Sedimentary nonmetallics (the phosphates, potash, sulfur, nitrates, and limestone) as well as bauxite and uranium (carnotite) also are associated with sedimentary rocks.

Associated with the igneous and metamorphic rock areas are most metals, for example, iron (usually), lead, copper, tin, the ferroalloys, gold, and silver. Most gems and some nonmetallic minerals (mica and asbestos) are found in the same associations. Uranium (pitchblende) occurs in these rocks.

Whereas these associations are well recognized, the mineral deposits themselves have a highly erratic geographical occurrence owing to the great variety of processes involving both mineral enrichment and dispersal which have occurred. The broad global pattern of rock classes with which minerals are related is delineated on the map in the illustration.

Employed and potential resources. Resources have meaning insofar as they are placed in use or are available for future exploitation. Distinction must be made between the employed and the potential resources. In practice this distinction is complex, but here only the simple geographical distinction will be noted. Employed resources are those which are significant to the present support of humans, at least locally. In general, the denser the population and the more advanced the technical arts of an area, the greater the need for production from resources, and employed resources become more nearly synonymous with all known resources. Thus, the recognized resources of the European peninsula and of the northeastern United States are mainly employed resources. On the other hand, the natural resources of the Amazon Basin, or much of Africa, Siberia are still largely potential resources.

Marine resources. Although the physical and biotic geography of the oceans is much less fully explored than that of the continents, enough is known to indicate that both living and mineral resources extend far beyond those presently exploited. The employed resources are rather sharply localized. The principal exploitation of marine animals and vegetation occurs: over the continental shelves; in the vicinity of the mixing of warm and cold currents; near large upwellings which occur particularly off the west coast of continents in lower middle latitudes; and adjacent to densely populated countries. Thus, the North Atlantic near Europe and from New England to Newfoundland contains heavily exploited fishing grounds, as do the seas near Japan, Korea, and southern California, and also waters of the Gulf of Mexico.

Minerals are derived from three separate types of marine sources: from sedimentary deposits underlying the continental shelves; from inshore deposits on the surface of the continental shelves; and from seawater. By far the most valuable of the mineral resources exploited from marine environments is petroleum; there has been an expansion of exploration, drilling, and pumping from beneath continental shelf waters, as off the Gulf coast of Louisiana, in the middle of the North Sea between Scotland and Norway, and offshore in the Persian or Arabian Gulf. Offshore placer deposits on the surface of the continental shelves yield gold, platinum, and tin. Common salt, magnesium, and bromine are derived directly from seawater. *See* OIL AND GAS, OFFSHORE.

Potential marine resources include the population of life-forms now exploited in some parts of the world, but not in others; animal and vegetative species now unused; fresh water from desalted seawater; and minerals so far unexploited which are in solution, are precipitated to the ocean bottom, or lie within rock below the surface of the continental shelf. One of the interesting speculative resources appears in the large quantities of so-called manganese nodules that cover some sections of sea bottoms at intermediate depths.

Resources of the continents. The resource pattern of the Earth may be summarized in a brief description of that for each continent and its neighboring waters.

Eurasian continent. As the largest landmass in the world, the Eurasian continent has the largest area of agricultural land in use, a very extensive total forest land area in which the softwood coniferous forest belt from Scandinavia to eastern Siberia predominates, and a wide variety of mineral resources. Great differences mark the major sections of the continent. The most productive agricultural areas are generally near the edge of the landmass, in western and central Europe and extending eastward into much of the central and southern sections of Russia in Europe; in the Indian subcontinent; and in mainland east Asia from the Red River valley in Vietnam northward to the Great Wall of China. The aggregate mineral endowment is outstanding and includes the largest known aggregate iron ore reserves in the world within Russia, located particularly in districts adjacent to the Urals; very substantial coal deposits, including the Ruhr field and the extensive coal beds in northern China and Manchuria; and what increasingly appears to be one of the two great concentrations of petroleum fields on Earth, namely, the Persian Gulf fields

shared by a number of separate states in the Middle East. The southeastern and eastern borders of the Eurasian heartland, moreover, have some of the great, but still undeveloped, hydroelectric generation sites of the world. Off the coasts of western Europe and Japan are the two most productively employed fisheries of the world. *See* ASIA; EUROPE.

Africa and Australia. Much of the entire area of Africa north of approximately 12°N must be classed as desert or semiarid, with few exotic water sources other than the Nile. Much of the remainder of the continent has wet-and-dry or rainforest climates, with the former predominating greatly in area. Seasonal drought, soil infertility, and widespread problems of laterization handicap agricultural exploitation. The east African highlands from Ethiopia southward, the high veld in South Africa, and the loftier sections of Zimbabwe, Rhodesia, and the Nile Valley and Delta are noteworthy exceptions. Except along the Nile there are still potential agricultural land resources, but they are comparatively minor. Associated with the extensive areas of igneous and metamorphic rocks which underlie much of the continent and particularly its southern half are outstanding deposits of metalliferous minerals, such as the copper ores astride the Zaire-Zambia boundary, the chrome-bearing ores of Zimbabwe, Rhodesia, and the gold deposits of the Witwatersrand in South Africa. Other mineral resources include diamonds, uranium, and, in Nigeria and the far north of the continent (Libya, Algeria), petroleum and natural gas. The water resources of mid-Africa include the largest potential hydroelectric power on Earth.

Similar general remarks may be made about Australia, whose much smaller area is covered mostly by desert, semiarid, and tropical wet-and-dry environments. Most of the agricultural productivity is peripheral, especially in the southeast. Metallic minerals at currently exploitable levels of size and richness support a substantial number of mining operations on the continent. *See* AFRICA; AUSTRALIA.

South America. The land resource is dominated by the unbroken extent of rainforest and wet-and-dry climates stretching east of the Andes from Colombia to northern Argentina and by substantial areas of water-deficient territory along the west coast and in the south and northeast of the South American continent. Some highly fertile soils in flat humid subtropical lands both west and east of the Paraná–La Plata river system are of minor extent by comparison with the whole. The Amazon Basin contains the largest stand of tropical hardwood forest on Earth. Metallic mineral resources are abundant in three general regions: the Andes, the largely crystalline rock highlands of eastern and southeastern Brazil, and the low plateau south of the lower Orinoco River. The Caribbean coast of Venezuela has the most productive petroleum reserves on the continent. *See* SOUTH AMERICA.

North America. Large sections of North American lands benefit from the advantages which characterize midlatitude humid-land resources under present technology. Disadvantages of desert and semiarid environments in much of the western half of the North American continent are tempered somewhat by the interspersal of mountain ranges throughout these drier regions. Taiga and other northern climatic environments are in considerable part coincident with the igneous and metamorphic rocks of the Laurentian Shield; and tropical environments are of small extent. In sum, this continent may be considered to have one of the best-balanced sets of resources, considering its substantial endowment in minerals of many different kinds, extensive forest lands, large annually renewed fresh-water supplies, great and varied agricultural lands, and the productive fisheries off both Atlantic and Pacific coasts. In addition, evidence has accumulated that in the southern section of the continent there may exist a major concentration of petroleum and natural gas resources extending southeastward from eastern Mexico and the contiguous continental shelf beneath the waters of the Gulf of Mexico to the fringes of the western Caribbean, with eventual productivity on the order of magnitude of the cluster of fields in the Middle East. Finally, North America has the highest ratio of employed resources to land area of all continents. In addition, it still contains considerable potential resources, including vast hydroelectric potential in eastern Canada (Quebec), and extensive (mostly) untapped oil deposits in the shale oil beds of Colorado and Wyoming, and the tar sands (thick oil in sandstone) of Alberta. *See* NORTH AMERICA.

Summary comment. The Earth's resource pattern has certain general characteristics. (1) Minerals usable under present technology are found in every environment, although mineral types differ according to location in sedimentary or igneous and metamorphic rock areas. Mineral exploration will continue indefinitely in all land areas, but the mineral resource possibilities of North America and the European part of the Eurasian continent have been examined in greater detail than those of any other large area. Ocean basins are the least-known part of the world as to mineral possibilities. (2) Agricultural lands and forest lands usable under present technology are dominated by those lying in midlatitudes. Sections of the taiga are important as forest resources. (3) The great potential agricultural and forest resources, if some technological improvement is assumed, lie within the humid tropical environments.

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Earth rotation and orbital motion

The rotation of the Earth about its axis is demonstrated by the classical Foucault pendulum experiment. Its revolution in its orbit around the Sun is shown by the annual parallactic displacement of

relatively nearby stars against the background of more distant stars. However, because the Earth is not truly a rigid symmetric body and because it interacts with other members of the solar system gravitationally, these motions vary with time. *See* FOUCAULT PENDULUM; ORBITAL MOTION; PARALLAX (ASTRONOMY).

Rotation of the Earth

Historically the rotation of the Earth has been an essential part of timekeeping. It was assumed that the rotational speed of the Earth was essentially constant and repeatable, and that the length of the day resulting from this constant rotational speed was naturally useful as a measure of the passage of time. Astronomical observations, however, have shown that the speed with which the Earth is rotating is not constant with time. The variations in rotational speed may be classified into three types: secular, irregular, and periodic. The secular variation of the rotational speed refers to the apparently linear increase in the length of the day due chiefly to tidal friction. This effect causes slowing of the Earth's rotational speed and lengthening of the day by about 0.0005 to 0.0035 s per century.

The irregular changes in speed appear to be correlated with physical processes occurring on or within the Earth. These have caused the length of the day to vary by as much as 0.01 s over the past 200 years. Irregular changes consist of so-called decade fluctuations with characteristic periods of 5–10 years as well as variations that occur at shorter time scales. The decade fluctuations are apparently related to processes occurring within the Earth. The higher-frequency variations are now known to be largely related to the changes in the total angular momentum of the atmosphere. *See* ATMOSPHERIC GENERAL CIRCULATION.

Periodic variations are associated with periodically repeatable physical processes affecting the Earth. Tides raised in the solid Earth by the Moon and the Sun produce periodic variations in the length of the day of the order of 0.0005 s with periods of 1 year, $^1/_2$ year, 27.55 days, and 13.66 days. Seasonal changes in global weather patterns occurring with approximately annual and semiannual periods also cause variations in the length of the day of this order. *See* EARTH TIDES.

Knowledge of the rotational speed of the Earth is required for observers on the Earth who want to know the orientation of the Earth in an inertial reference frame. This includes navigators, astronomers, and geodesists. The rotational speed of the Earth remains essentially unpredictable in nature due to the incompletely understood irregular variations. Because of this, astronomical observations continue to be made regularly with improving accuracy, and the resulting data are the subject of continuing research in the field.

Observations of rotational speed. Astronomical observations of quasars, the Moon, and artificial Earth satellites are used to create a nonuniform time scale that is based strictly on the rotation of the Earth within an inertial reference system defined by the positions and motions of the celestial objects. This Universal Time scale (UT1) is compared with time scales known to be more uniform such as those determined from atomic clocks (International Atomic Time), the motion of the Earth about the Sun (Dynamical Time), or the motion of solar system objects. Variations in the differences among these types of time scales may be used to determine variations in the rotational speed of the Earth. Astronomical observations of time are made routinely by observatories located around the world for this purpose. Differences between the astronomically determined time scale and a uniform time scale are published routinely by the International Earth Rotation and Reference System Service (IERS). One-second adjustments are made at infrequent intervals to the uniform time scale Coordinated Universal Time (UTC) so that the difference between UT1 and UTC is never greater than 0.9 s. These adjustments are called leap seconds. UTC serves as the basis for civil time in most of the countries of the world. *See* ATOMIC CLOCK; ATOMIC TIME; DYNAMICAL TIME; TIME.

The times and locations of past total solar eclipse observations have also been analyzed to provide information on the length of the day. If the Earth has changed its speed of rotation since ancient times, the path of an eclipse that occurred thousands of years ago would be displaced in longitude with respect to the path that would have occurred if the rotational speed had remained constant. Ancient records of eclipses, while not made with great accuracy, are valuable because they were made long ago. Comparison of very old observations of the longitude of the Sun with current theories of the motion of the Sun based on a uniform time scale has also been used to estimate the increase in the length of the day since ancient times. *See* ECLIPSE.

Careful analyses of this information reveal the three types of variations in the speed of rotation. These ancient observational data form the basis for estimates of the secular deceleration in the speed of rotation. The more recent information, having been obtained with higher accuracy and more regularity, has shown the changes in the acceleration causing the irregular variations in the length of the day. These data have also been used to detect the periodic variations in the length of the day (**Fig. 1**).

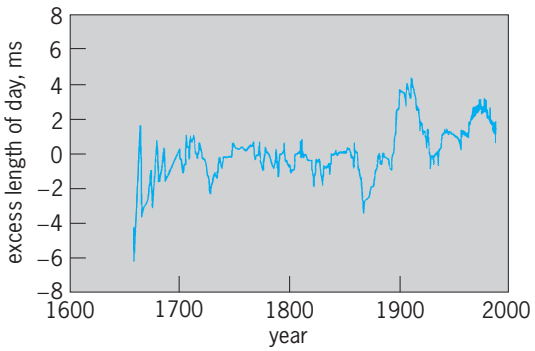

Fig. 1. Excess length of day (length of day—86,400 s), measured in milliseconds, since 1650.

Causes of variations. The conservation of angular momentum of the Earth requires that changes in the Earth's moments of inertia produce changes in the speed with which the Earth is rotating. The moments of inertia depend on the distribution of mass on and within the Earth. This includes the mass contained in the atmosphere and in the oceans. As mass is redistributed, the moments of inertia change, producing changes in the rotational speed of the Earth. An example of such an effect is the spinning ice skater who is able to change rotational speed by redistributing mass by changing the positions of the arms. *See* ANGULAR MOMENTUM; MOMENT OF INERTIA; ROTATIONAL MOTION.

The crust and the interior of the Earth are not strictly rigid. Tides generated by the Sun and the Moon change the shape of the Earth. The Earth is thought to have a fluid core with material possibly in motion. Coupling between such a core and the mantle would change the speed of rotation of the mantle from which the observations are made. The Earth may undergo changes in the moments of inertia as landmasses move or rise and fall. All of these motions cause changes in the Earth's moments of inertia and may cause changes in the rotational speed. Seasonal and short-term variations are caused by an exchange of momentum between winds and the crust of the Earth. Precise calculations of the total angular momentum using global observations of the wind speed and direction demonstrate this effect. The variations in the rotational speed are apparently caused by the interaction of the winds with the Earth's topography, and may be the leading cause for most of the irregular high-frequency variations in the rotation of the Earth.

Another known cause of variations in the rotational speed is tidal friction. The Moon raises tides in the ocean. Friction carries the maximum tide ahead of the line joining the center of the Earth and Moon (**Fig. 2**). The resulting couple diminishes the speed of rotation of the Earth, and this reacts on the Moon to increase its orbital momentum. The sum of the angular momentum of the Earth and the orbital momentum of the Moon remains constant. This produces an increase in the size of the orbit of the Moon and a reduction of its angular speed about the Earth. Tidal friction should be distinguished from actual changes in the moments of inertia of the Earth brought about by the tides. *See* TIDE.

The effect of tidal friction is to increase both the distance of the Moon and the length of the lunar month measured with a uniform time scale. Because of the change in the lunar month, the Moon is observed to have an orbital deceleration in terms of a uniform time scale. The proportional change in the distance, however, is greater than in the length of the month. Hence, in terms of Universal Time, the Moon appears to have a secular orbital acceleration, an effect discovered by E. Halley in 1693 from a study of ancient eclipses. In 1853 J. C. E. Adams found that gravitational theory could account for only about half of the acceleration found by Halley. It has become clear since then that the rotational speed is also decreased because of tidal friction.

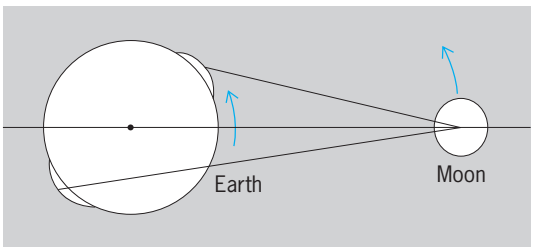

Fig. 2. Couple produced by tidal friction.

Revolution about the Sun

The motion of the Earth about the Sun is seen as an apparent annual motion of the Sun along the ecliptic. That the effect is caused by the motion of the Earth and not that of the Sun is proved by the annual parallactic displacement of nearby stars and by the aberration of light, which causes an apparent annual displacement of all stars on the celestial sphere. *See* ABERRATION (ASTRONOMY).

Orbit of the Earth. A large number of astronomical observations of the positions of the Sun and other solar system objects have been made and are being made continuously. This information is required to determine the nature of the motion of the Earth about the Sun. Observations are analyzed using the mathematical methods of celestial mechanics to provide improved estimates of the motions of the solar system objects in the future and to describe the past motions of the objects. The description of the apparent motion of the Sun in the sky provides the determination of the orbit of the Earth. The Earth moves around the Sun in a plane called the ecliptic. The orbit is characterized in part by the period of revolution, the mean radius, and its eccentricity, a quantity that describes the elliptical nature of the orbit. *See* CELESTIAL MECHANICS; ECLIPTIC; ORBITAL MOTION.

Period of revolution. The true period of the revolution of the Earth around the Sun is determined by the time interval between successive returns of the Sun to the direction of the same star. This interval is the sidereal year of 365 days 6 h 9 min 9.51 s of mean solar time or 365.25636 mean solar days. The period between successive returns to the moving vernal equinox is known as the tropical year of 365 days 5 h 48 min 45.2 s or 365.24219 days. The length of the tropical year is regarded as the length of the year in common usage for calendars. The period of time between successive passages at perihelion (the closest approach of the Earth to the Sun) is called the anomalistic year of 365 days 6 h 13 min 53.26 s or 365.25964 days. The length of each of these years depends on observationally determined astronomical quantities. Improvements in the determination of these quantities will result in slight changes in the numerical values. The lengths of the years listed above are given for the year 2000. These values vary slowly as a consequence of the long-period perturbations of the Earth's orbit by other planets. *See* CALENDAR; PERTURBATION (ASTRONOMY).

Mean radius of orbit. The mean distance from the Earth to the Sun, or the semimajor axis of the Earth's orbit, was the original definition of the astronomical unit (AU) of distance in the solar system. Its absolute value fixes the scale of the solar system and the whole universe in terms of terrestrial standards of length. The distance between the Earth and the Sun can be determined by a variety of methods; one method is measuring the solar parallax. The Sun's mean equatorial horizontal parallax p is the angle subtended by the equatorial radius r of the Earth at the mean distance a of the Sun. Mathematically, $a = r/\sin p = 206,265\ r/p$, if p is in seconds of arc. The equatorial radius of the Earth is 3963.182 mi. (6,378,136 m). Geometrical, gravitational, and physical methods have been used at various times to measure the solar parallax. *See* ASTRONOMICAL UNIT; PARALLAX (ASTRONOMY).

Geometrical methods. The geometrical methods involve the direct measurement by optical triangulation of the parallax of a nearby planet (Mars or Venus) or other nearby solar system object such as the minor planet Eros at its closest approach to the Earth. Because the relative distances in the solar system are accurately known in terms of the astronomical unit, the absolute measurement of one very accurate distance can be used to calibrate the scale of the system (**Fig. 3**).

Gravitational methods. The gravitational methods involve the determination of the ratio of the mass m of the Earth to the mass M of the Sun from the perturbations in the motion of a minor planet (such as Eros) caused by the Earth. The method also depends essentially on an accurate determination of the acceleration of gravity g and the length of the sidereal year T in seconds of mean solar time. The distance a is then found from the application of Kepler's third law, $4\pi^2a^3/T^2 = gr^2[1 + (M/m)]$. Because g and T are known with high accuracy, a determination of M/m gives a directly in terms of terrestrial standards of length. Although it is customary to express this result as a parallax angle, the Earth's radius is not needed to compute a, but it is necessary to make allowances for the mass of the Moon relative to the mass of the Earth. *See* GRAVITY.

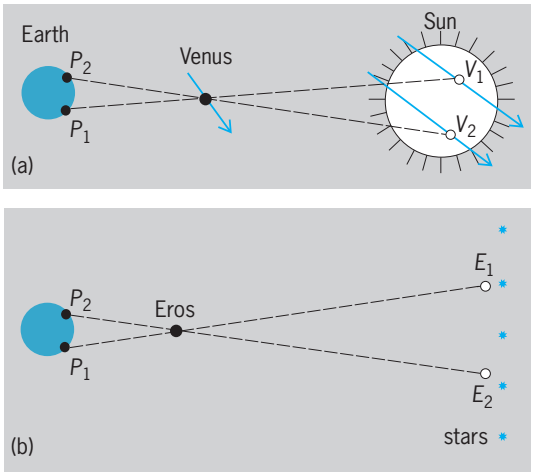

Fig. 3. Determination of solar parallax. (*a*) Observation of Venus in transit. (*b*) Observation of Eros in opposition.

Physical methods. One physical method depends on a determination of the ratio of the mean orbital velocity, $V = 2\pi a/T$, to the accurately known velocity of light. This ratio can be derived either from the annual variation of the radial velocities of ecliptic stars (or occasionally planets) determined by observations of the Doppler shift of spectral lines, or with less accuracy, from the constant of aberration. *See* DOPPLER EFFECT.

The most precise method relies on measurement of the travel time of electromagnetic signals reflected from objects in the solar system or received from artificial interplanetary probes. The distance of the planet is the ratio of half of the round-trip travel time of the electromagnetic signal to the velocity of the electromagnetic waves. Since the distance of the planet is known precisely in astronomical units from the orbital ephemerides, measuring the actual distance in meters at some given time permits the astronomical unit to be calibrated in terms of terrestrial units of length. Also the Doppler shift of radar reflections gives the radial velocity in $m \cdot s^{-1}$, which, on comparison with the known value in $AU \cdot s^{-1}$, gives the astronomical unit in meters. The currently adopted value of the astronomical unit is $1.495978706 \times 10^{11}$ m (92,955,807 mi).

Eccentricity of orbit. The eccentricity of the Earth's orbit can be determined by the variations of the apparent diameter of the Sun's disk. Determinations of higher accuracy are based on the variable speed of the Sun's apparent motion along the ecliptic and the laws of elliptic motion. The nonuniformity of the Sun's motion manifests itself in the equation of time, which is the difference between solar time determined from the actual observation of the Sun, and mean solar time which is based on the motion of a conventionally defined fictitious point having a uniform motion close to the average of the Sun's motion. The difference in the two types of time arises in part from the obliquity of the ecliptic (the angle between the equatorial plane of the Earth and the plane of the ecliptic) and in part from the eccentricity of the Earth's orbit. The adopted value of the eccentricity is 0.0167086171540. The number varies slowly with time due to perturbations from the planets. The Earth is at perihelion (closest distance to the Sun) on January 2. It reaches its greatest distance from the Sun on July 2. *See* EQUATION OF TIME.

Seasons. The fact that the Equator of the Earth is inclined in space by about 23.°5 to the orbital plane of the Earth (the ecliptic) causes the Northern Hemisphere to be exposed to the more direct rays of the Sun during part of the Earth's revolution around the Sun. The Southern Hemisphere receives the more direct rays 6 months, or a half revolution, later. This effect causes the seasons.

Other Motions

In addition to the rotation of the Earth and its orbital motion about the Sun, the Earth experiences various

small motions about its center of mass. Precession and nutation are examples, and these are caused by the gravitational attraction of the Sun and Moon on the nonspherical Earth. Because the Earth is ellipsoidal in shape, the gravitational attraction of these bodies produces a couple acting on the equatorial bulge, changing the orientation of the Earth about its center of mass. This motion can be predicted with accuracy, based on the knowledge of the shape of the Earth and of the motion of the Earth around the Sun and that of the Moon about the Earth. Precession causes the axis of angular momentum of the Earth to describe a 23.°5 cone in space with a period of about 26,000 years. Nutation causes the axis of angular momentum to "nod" slightly in space as it executes the precessional motion. The main period of the nutational motion is 18.6 years. Other periodic motions due to the gravitational attraction of the Sun and the Moon are also included in nutation. *See* NUTATION (ASTRONOMY AND MECHANICS); PRECESSION OF EQUINOXES.

Because the axis of symmetry of the Earth is not aligned precisely with the axis of rotation, the Earth also executes a motion about its center of mass known as polar motion. This motion, caused by geophysical and meteorological effects on and within the Earth, can be detected by the motion of the Earth's axis over the surface of the Earth. It is not predictable with accuracy, and must be observed continuously to provide the most precise information on the orientation of the Earth. Polar motion is characterized mainly by approximately 435-day and a 365-day periodic quasicircular motions of the axis of rotation on the surface of the Earth. The radius of the circular motion is of the order of 16 ft (5 m), but this varies. *See* EARTH; PLANET. Dennis D. McCarthy

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Earth sciences

Sciences that involve attempts to understand the nature, origin, evolution, and behavior of the Earth or of its parts and to comprehend its place in the universe, especially in the solar system. Understanding has advanced primarily through improved appreciation of the complex, usually cyclical interactions that take place among distinct parts of the Earth such as the lithosphere, atmosphere, hydrosphere, and biosphere. Geophysics is the study of the physics of the Earth, emphasizing its physical structure and dynamics. Geochemistry is the study of the chemistry of the Earth, dealing with its composition and chemical change. Geology is the study of the solid Earth and of the processes that have formed and modified it throughout its 4.5-billion-year history. *See* GEOCHEMISTRY; GEODESY; GEOLOGY; GEOPHYSICS; SOLAR SYSTEM.

Many branches of geology are considered separate sciences. Mineralogy is the study of the composition, structure, and properties of minerals. Petrology involves understanding how rocks originate and evolve, as well as rock description and classification. Specialties related to petrology include sedimentology and volcanology. Stratigraphy is the study of the origin, age, and development of layered, generally sedimentary rocks. Paleontology is the study of ancient (fossil) life. Historical geology is the study of the evolution of the Earth and its life. Geomorphology is the study of landscapes and their evolution. Seismology is the study of earthquakes and their effects. Structural geology is the study of deformed rocks. Engineering geology relates to the support of human constructions by underlying rock. *See* ENGINEERING GEOLOGY; GEOLOGY; GEOMORPHOLOGY; HYDROLOGY; MINERALOGY; PALEONTOLOGY; PETROGRAPHY; PETROLOGY; SEISMOLOGY; STRATIGRAPHY; STRUCTURAL GEOLOGY; VOLCANOLOGY.

Oceanography is the study of the oceans; limnology, the study of lakes; hydrology, the study of underground and surface water; and glaciology, the study of glaciers, ice caps, and ice sheets. These disciplines address the study of water in and on the Earth. The gaseous outer parts of the planet are the province of the atmospheric sciences, including meteorology, which is concerned with the weather and weather forecasting; climatology, which deals with longer-term and regional variations; and aeronomy which, because it deals with the outermost ionized region of the atmosphere, is much concerned with solar terrestrial interactions, including the aurora borealis and aurora australis. The biosphere embodies all life on Earth, and its study includes molecular biology, zoology, botany, and ecology. Geography, the study of all that happens at the Earth's surface, has been distinct insofar as it has encompassed not only physical and biological sciences but also the social sciences, including aspects of political science and economics. This distinction is fading rapidly as other earth sciences become more involved with social considerations.

A great deal of overlap exists among the earth sciences, and all have been subdivided into many disciplines. But progressively, especially since the International Geophysical Year (IGY) in 1957, subdivision has been complemented by integration as scientists have become more aware of the ways in which parts of the Earth system can interact. In trying to assess any possible impacts of predicted global warming, scientists have discerned interactions attributable to the burning of fossil fuels and the destruction of forests over the past 200 years. In the twenty-first century the effects of those interactions

could lead to both rising atmospheric temperatures and rising sea level. The prospect of possible adverse changes in the human environment has generated concern, and for that reason earth scientists are working with politicians to clarify what is happening and, as deemed necessary, to help initiate appropriate responses.

There are two main reasons why earth scientists have come to consider that they can help address such issues. First, Earth-observing instruments on satellites in geostationary and low Earth orbit now permit the acquisition, processing, and storage of the very large data sets that are required to show how the atmosphere, ocean, biosphere, hydrosphere, and lithosphere are interacting and changing globally on time scales from daily to decadal. Second, advanced computational capabilities now allow scientists to manipulate the large, satellite-derived data sets and also to produce models of the behavior of the interactive systems. These activities are in their infancy. It is not possible to say how they may develop as the earth sciences evolve. The success of scientists in predicting, months in advance, the advent of a severe ENSO (El Niño Southern Oscillation) is an example of the kind of activity that may become more common. One such ENSO prediction was far enough in advance to enable farmers to switch to the planting of drought-resistant crops.

Although studies in the earth sciences that are focused on the well-being of the growing human population are likely to be increasingly important, basic understanding in the earth sciences is also likely to come to depend more on very large data sets. Improved understanding of the fundamentals of the Earth system will require the acquisition of very large data sets, their manipulation, processing, and storage. This will be complemented by the improving ability to construct models of the parts of the Earth system and of the ways in which those parts interact. Models will be constructed, tested, and abandoned in favor of better models as more and better data are acquired. Quality control on both data acquisition and processing will be critical. Kevin Burke

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Earth tides

Cyclic motions of the Earth, sometimes over a foot or so in height, depending on latitude, caused by the same lunar and solar forces which produce tides in the sea. These forces also react on the Moon and Sun, and thus are significant in astronomy in evaluations of the dynamics of the three bodies. For example, the secular spin-down of the Earth due to lunar tidal torques is best computed from the observed acceleration of the Moon's orbital velocity. In oceanography, earth tides and ocean tides are very closely related. *See* EARTH, GRAVITY FIELD OF; GEODESY; TIDE.

By far the most widely used earth tide instruments are the tiltmeter and the gravimeter. Both instruments have the merits of portability, high potential precision, and low cost. Thus they are able to advance economically an important mission—the global mapping of earth tides and ocean tides. *See* GRAVITY METER; SEISMOGRAPHIC INSTRUMENTATION.

Tide-producing potential. Tidal theory begins simply, by evaluating the tidal forces produced on a rigid, unyielding, oceanless, spherical Earth by the Moon and Sun. The Earth, Moon, and Sun can be regarded as perfectly spherical; the results of this idealized assumption will be modified later in this text to relate earth tide observations to deformable Earth models.

In the expansion of the total gravitational potential of the satellite as a sum of solid spherical harmonics, the tide-producing potential consists only of those terms which vary within the Earth and are the source of Earth deformations. The omitted terms are the constant which is arbitrarily chosen to produce null potential at the Earth's center and the first-degree term expressing the uniform acceleration of every part of the Earth toward the satellite (thus also entailing no distortion). Accordingly, the tide-producing potential U is expressible as the sum of solid spherical harmonics of degree 2 and higher in the notation which follows.

If the mass of the satellite is denoted by m, its distance from the Earth's center by R, the distance of the observing point from the Earth's center by r, and the geocentric zenith angle measured from the line of centers by θ, then for points $r < R$, Eq. (1)

$$U = U_2 + U_3 + \cdots = GmR^{-1}\left[\left(\frac{r}{R}\right)^2 P_2(\cos\theta) + \left(\frac{r}{R}\right)^3 P_3(\cos\theta) + \cdots\right] \quad (1)$$

holds, where G is Newton's universal gravitational constant (6.67×10^{-8} cm^3 g^{-1} s^{-2}).

On an idealized oblate Earth of equatorial radius r_e and flattening constant f, a point on the surface at geocentric latitude ϕ has radial distance $r = r_e C(\phi)$, where $C(\phi) \equiv 1 - f \sin^2\phi$. The vertical (upward) component of tidal gravity $\partial U/\partial r$ and the horizontal component $r^{-1}\partial U/\partial\theta$ in the azimuth direction away from the satellite, with $P^1{}_n(\cos\theta) = -\partial P_N(\cos\theta)/\partial\theta$, are given by Eqs. (2) and (3). Here

$$g_r = \frac{\partial U}{\partial r} = GmC\alpha^3 r_e^{-2} \sum_{n=2}^{\infty} n(\alpha C)^{n-2} P_n(\cos\theta) \quad (2)$$

$$g_\theta = r^{-1}\frac{\partial U}{\partial\theta} = -GmC\alpha^3 r_e^{-2} \sum_{n=2}^{\infty} (\alpha C)^{n-2} P_n^1(\cos\theta) \quad (3)$$

$\alpha_m \equiv r_e/R_m$ and $\alpha_S \equiv r_e/R_S$ for the Moon and Sun, respectively.

TABLE 1. Parameters for some equilibrium tides

Species	Darwin symbol	Doodson number	Period	Amplitude factor, *b*
Long period	Lunar	055565	18.6 y	0.066
	Sa	056554	1 y	0.012
	SSa	057555	$^1/_2$y	0.073
	MSm	063655	31.85 d	0.016
	Mm	065455	27.55 d	0.083
	MSf	073555	14.77 d	0.014
	Mf	075555	13.66 d	0.156
	—		13.63 d	0.065
Diurnal	O_1	145555	25.82 h	0.377
	P_1	163555	24.07 h	0.176
	K_1	165555	23.93 h	0.531
Semidiurnal	N_2	245655	12.66 h	0.174
	M_2	255555	12.42 h	0.908
	S_2	273555	12.00 h	0.423
	K_2	275555	11.97 h	0.115

Harmonic constituents. Equations (2) and (3) indicate the simple nature of the dependence of tidal gravity on the rigid Earth upon the distance R and the zenith angle θ of the satellite. But the time variations of R and θ are complex, because of the Earth's rotation and the complexity of the orbital motions. The objective in earth tide observations is assistance in determining present relevant properties of the Earth. **Table 1** lists the periods and amplitude factors b of the larger tidal harmonic constituents over the broad range of tidal periods. In classifying tidal constituents mnemonically, G. H. Darwin used alphabets in terms of pairs with respect to the Moon (M) and the Sun (S), and A. T. Doodson used argument numbers, which are the coefficients of the six astronomical, independent variables arranged in order of increasing speed. The first three argument numbers are called the constituent number, of which the first two-digit number is the group number. The first digit of the group number designates species 0, 1, and 2 for long-period, diurnal, and semidiurnal, respectively. The purely geometric complexities essential in applied tidal theory are deemphasized here and, when feasible, the simple zenith-centered satellite coordinate system used in Eqs. (1)–(3) is adopted.

Tidal torques. The orbital acceleration $\dot{n}_m$ of the Moon is the critical quantity in Eq. (4), giving the

$$N_m = -^1/_3[M_e M_m/(M_e + M_m)]r_m^2 \dot{n}_m \qquad (4)$$

lunar torque N_m, which slows the Earth's spin. (Here M_e and M_m are masses of the Earth and Moon respectively, and r_m is the distance to the Moon.) The value for N_m has been calculated to equal -3.9×10^{23} dyne-cm, to which corresponds the energy loss rate $\dot{E}$, shown in Eq. (5), and the relative spin-down,

$$\dot{E} = N(\omega - n_m) = 2.7 \times 10^{19} \text{ ergs/s} \qquad (5)$$

$-\dot{\omega}/\omega = N/c\omega = 0.21$ eon^{-1} (1 eon $= 10^9$ years). Here ω is the Earth's rotational velocity, and c its principal moment of inertia.

There is an observation equation for the M_2 ocean tidal constituent from which an estimate of the tidal acceleration of lunar longitude is made; the value $\dot{n}_m = -27.4'' \pm 3.0''$ century^{-2} is obtained. That means $N_m = -4.8 \times 10^{23}$ dyne-cm, to which corresponds the energy loss rate $E = 3.3 \times 10^{19}$ ergs/s, and the relative spin-down value of 0.26 eon^{-1}.

Tidal loss in the solid Earth. The gravest free mode of the Earth, period 0.9 h, has the same external form and nearly the same internal geometry as the M_2 tide bulge. Its observed $Q = 350 \pm 100$ can be used to estimate the loss rate $\dot{E}_B$ in the bodily M_2 tide (Q^{-1} is the loss rate number, and M_2 is the principal semidiurnal tide). The result, $\dot{E}_B = 7 \times 10^{17}$ ergs/s, is only 3% of the required total in Eq. (5). The rate at which the Earth dissipates the total lunar-solar tidal energy has also been estimated to be $5.7 \pm 0.5 \times 10^{19}$ ergs/s, the share attributed to the oceans is $5.0 \pm 0.3 \times 10^{19}$ ergs/s.

Instrumental dimensionless amplitude factors. The reading of an earth tide meter is altered by the fact that it is anchored to a yielding Earth. The Earth chosen is assumed to be the radially symmetric standard.

Gravimeter. Basically, a gravimeter consists of a mass generally supported by a spring. (The superconducting model used a magnetic field.) Variations in gravity are measured by the extension of the spring or in the null method by the small corrections required to restore the original configuration. This low-drift-rate superconducting gravimeter is designed to increase to several months the intervals between replenishing the helium supply. It is expected to give high precision in gravity measurements. A satellite affects the mean local value of gravity in three ways: by its direct attraction, by the tidal change in elevation of the observing station, and by the redistribution of mass in the deformed Earth.

Tiltmeters. The tiltmeter measures changes in the angle of tilt between the tiltmeter's foundation and the local vertical, both subject to tidal variations. Several types of tiltmeters are in use. In the horizontal pendulum the rotation axis is fixed at a small angle with the vertical to produce high sensitivity to tilts. In the level tube, which may be several hundred meters or more long, the difference in elevation of a fluid is measured at the two ends. In this way a sample of the tilt of a large region is obtained, but the

horizontal pendulum sometimes also samples a large volume.

Extensometer. The extensometer, or linear strain meter, measures the change in distance between two reference points. In the Benioff design these positions are connected with a quartz tube to bring the points into juxtaposition for precise measurements of relative displacement. By using laser beams, the measurements may be made over distances of kilometers, but the short-baseline strain-mechanical meters still have their use.

Indicated and true phase lags. The phase indications of a gravimeter or tiltmeter are only a fraction of the true angular lag of the tidal bulge. The tidal bulge is assumed to retain its no-loss equilibrium form, but with the axis carried forward from the Moon's zenith by the small angle ϵ required to produce the Moon's known orbital acceleration.

Theories and characteristic numbers. Three basic numbers, Love numbers h and k and Shida number l, characterize the elastic behavior of an elastically yielding Earth. Thus h represents the ratio of the tidal height of the yielding Earth to that of the equilibrium tide, k the ratio of the additional potential due to the tidal deformation of the yielding Earth to the deforming potential, and l the ratio of horizontal displacement of the yielding Earth to that of the equilibrium tide.

Static theory. Static theory of earth tides (the frequency domain) is formulated for a spherical, nonrotating, and isotropic Earth. It is based on the reduction of the problem of elastic deformation of a sphere to a system of six first-order differential equations which are subjected to numerical integration in one formulation. When these equations are integrated with different Earth models by means of the Runge-Kutta method, the second-order Love numbers that are derived are virtually independent of either the presence of the solid inner core or the presence or absence of a low-velocity layer in the upper mantle (**Table 2**). Also, the continental or oceanic crustal model is only of second order with respect to the effects of the oceans.

Dynamic theory. Theoretically, a torque-free, nearly diurnal nutation or resonance is indeed possible by the presence of the liquid core inertially coupled to the mantle, and the predicted motion can give a resonance amplification to the nearly diurnal tides. However, theoretical Earth models, while complicated, are a great simplification of the actual Earth.

Time domain. Traditionally, theories of tides for the Earth, the oceans, or the atmosphere were always in the frequency domain. However, tides as observed are typically in the time domain, that is, in time series. Although development of theories of tides in the time domain is more complicated than that in the frequency domain, for practical purposes it would be more realistic to develop theories of tides in the time domain. The Earth, the oceans, and the atmosphere eventually must be treated as a single system, responding to the complete spectrum of the tide-generating forces, wind stresses, and interactions, as well as other periodic and nonperiodic dynamic forcings.

The time-domain earth tide is solved for a spherical, nonrotating, elastic, and isotropic Earth, and also for a rotating Earth. In their formulation, both the spheroidal and toroidal components are simultaneously introduced into the general form of the linearized lagrangian equations of motion, which govern small elasto-gravitational disturbances away from equilibrium of an arbitrary, uniformly rotating, self-gravitating, elastic Earth with an arbitrary initial state of stress field. The resulting equations of motion are those of a set of integro-partial differential equations, the solutions to which satisfy a set of boundary conditions in the Earth. There are strong coupling effects due to the rotation of the Earth.

Interaction of earth and ocean tides. Neither the problem of earth tides nor that of ocean tides can be solved independently. Characterization of ocean tides in the open oceans is uncertain. Calculations of ocean tides in the open oceans made directly from Laplace's tidal equations without consideration of the elastic yielding of the Earth due to the tidal and load deformation are inadequate. The interdependency of the problems of earth and ocean tides is a problem in geodynamics.

Tidal gravity and ocean tides. A transcontinental tidal gravity profile, consisting of nine stations across the United States along the 40th parallel, show the consistent results of the dominant effect of the ocean tides in the Pacific and Atlantic oceans (**Figs. 1** and **2**).

Ocean tides are the primary cause of the observed tidal gravity variations. These perturbations essentially depend only on the amplitude and phase of the ocean tide and the distance between the

TABLE 2. Earth tide parameters for different Earth models

Model	h	k	l	k/h	δ	γ	l/h
Alsop and Kuo							
Gutenberg-Bullen	0.619	0.305	0.088	0.493	1.162	0.686	0.142
Jeffreys-Bullen B	0.618	0.304	0.088	0.492	1.162	0.686	0.142
Jeffreys-Bullen A	0.607	0.299	0.081	0.493	1.158	0.692	0.133
Gutenberg-Bullen A	0.608	0.300	0.082	0.493	1.158	0.692	0.135
Farrell							
Gutenberg-Bullen Earth model	0.3040	0.6114	0.0832	0.497	1.1554	0.6926	0.136
Oceanic mantle	0.3055	0.6149	0.0840	0.497	1.1567	0.6906	0.137
Shield mantle	0.3062	0.6169	0.0842	0.496	1.1576	0.6893	0.136

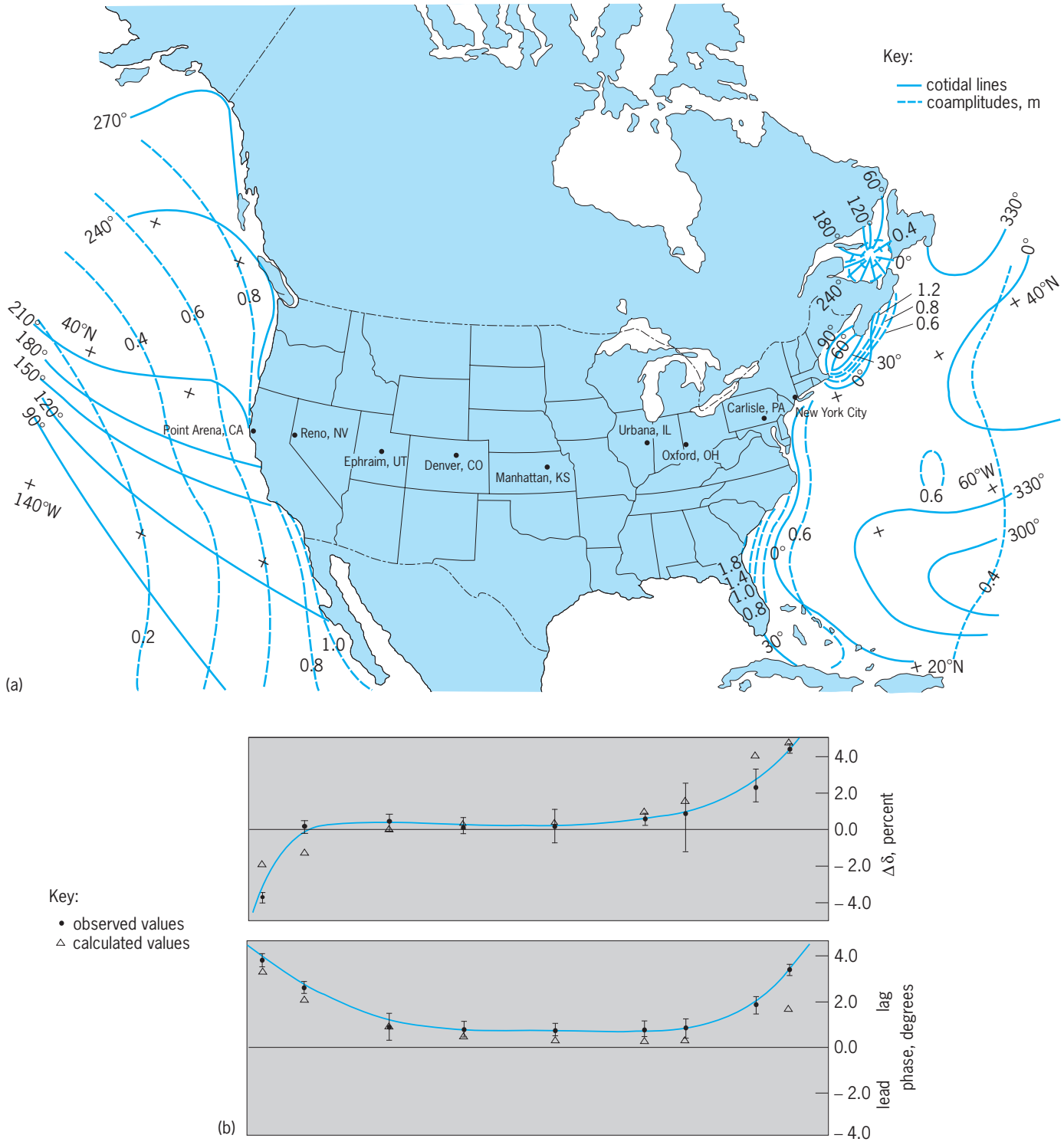
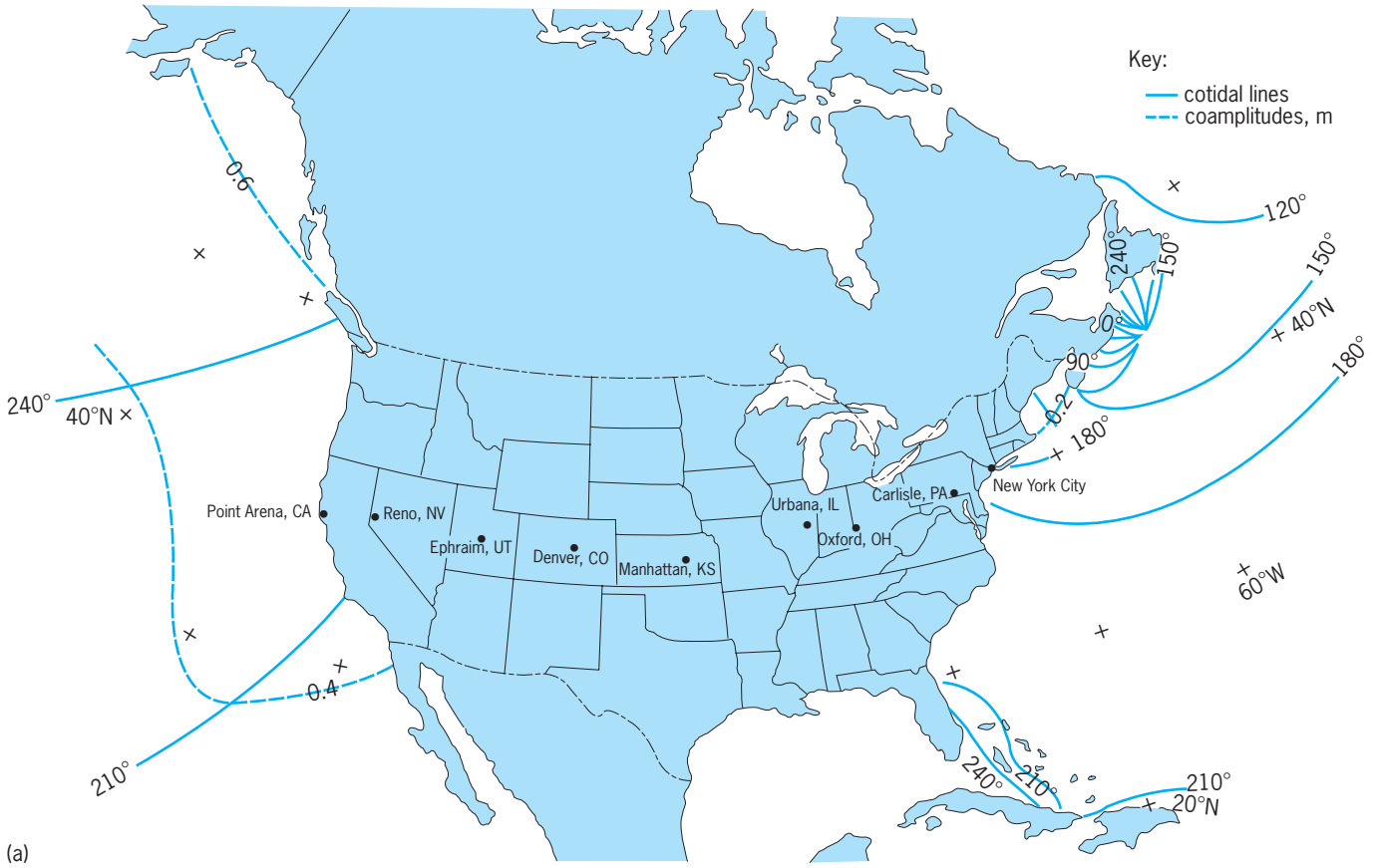


Fig. 1. Effect of the ocean tides in the Atlantic and Pacific oceans on the earth tides for the semidiurnal tidal constituent M_2. (*a*) Ocean tides. (*b*) Comparison of the observed and calculated values. (*After J. T. Kuo et al., Transcontinental gravity profile across the United States, Science, 168:968–971, 1970*)

observation station and the ocean tidal load; land-based and island-based tidal gravity measurements can provide an independent set of observation data. If tidal gravity measurements are made accurately, it is more appropriate to consider the inverse problem of mapping ocean tides in the open oceans with extended earth tidal gravity measurements on the adjacent lands, supplemented by ocean-bottom stations and the ocean tidal information derived from nearshore and island stations.

Open ocean tides can be mapped by solving the inverse problem using land- and island-based tidal gravity measurements, coupled with shore and island



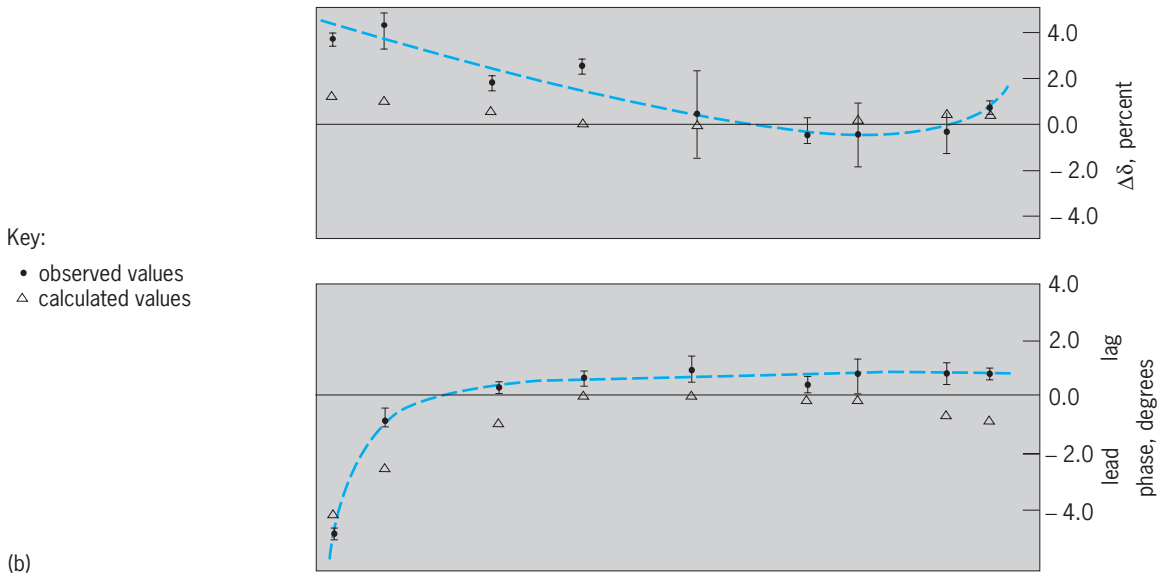


Fig. 2. Effect of the ocean tides in the Atlantic and Pacific oceans on the earth tides for the diurnal tidal constituent O_1. (*a*) Ocean tides. (*b*) Comparison of the observed and calculated values. (*After J. T. Kuo et al., Transcontinental gravity profile across the United States, Science, 168:968–971, 1970*)

ocean tidal measurements and a few ocean-bottom measurements.

Modified Laplace's tidal equations. Understanding of open ocean tides can be obtained both through numerical integration of Laplace's tidal equations and through direct measurements of tides in the deep oceans. However, the coamplitude and cophase tidal charts, principally of the tidal constituent M_2 produced by means of the numerical integration of Laplace's tidal equations, apparently are quite sensitive to the boundaries of the ocean basins and the law of friction, and generally fail to give close agreement with the tidal observations on mid-oceanic islands. Thus, the original formulation of Laplace's tidal equations, which are based on the assumption of a rigid Earth, are inadequate to serve as a theoretical basis

for calculating the ocean tides. The contribution of the tidal and loading deformation of the elastic Earth cannot be neglected in the formulation of the tidal equations.

Tidal tilt and tidal strain. Measurements of tidal tilt and tidal strain are more complicated than those of tidal gravity. Unlike tidal gravity measurements, which are virtually independent of inhomogeneities of the elastic properties in the Earth's crust or even in the upper mantle, tidal tilt and tidal strain measurements, in addition to the influence of ocean tidal loading, are very sensitive to the influences of the inhomogeneities, including the site of instrumental installation, topography, and geological structure.

It is well known that the underground openings in otherwise uniform rock mass strongly distort the elastic strain field. Tiltmeters installed in underground tunnels generally measure cross-tunnel strain-induced tilts, in addition to the regional tilt.

Moreover, the inhomogeneities in the Earth's crust, such as variable surface topography and geological discontinuities and the cavity in which the instruments are installed, are recognized as having significant influence in tidal tilt and tidal strain measurements. Tidal strain was observed from seven strain stations across the continental United States. A comparison of the tidal strain as predicted for a radially symmetrical, stratified, and homogeneous Earth model with corrections for the loading effects of the worldwide ocean tides (derived from the numerical integration of Laplace's tidal equations) indicated that the topographic and geological influences on tidal strain amount to as much as ±25% of the tidal strain for some of the strain stations, and the topographic effect on tidal strain exceeded ±50% of the tidal strain in some extreme cases.

Satellite and interferometry measurements. It is possible to estimate values for the amplitudes and phases of the M_2, K_1, and P_1 ocean tidal constituents. With improved accuracy of satellite orbital determination, long time spans of accumulated data would permit an improvement in the accuracy of ocean tidal parameters in the near future.

Very long baseline interferometry (VLBI) has also been used to determine vector separations between radio telescopes and positions of radio sources. It has a precision of under 2 in. (5 cm) for baselines from 600 to 2400 mi (1000 to 4000 km) and less than about 4 in. (subdecimeter) for intercontinental lengths of about 3600 mi (6000 km).

Nearly diurnal nutation. The motion of the nearly diurnal nutation (or resonance) describes the free oscillation of the mantle's inertia axis about the axis of the Earth's rotation. The natural period of such a resonance of the Earth depends on the internal structures and elastic properties of the Earth and is not exactly equal to a sidereal day. However, calculations indicate that only the major K_1 and the minor ψ_1 tidal constituents with a period close to a sidereal day have amplitudes changed by the resonance effect. The theoretical results of Molodensky model II have been compared with the observed values of

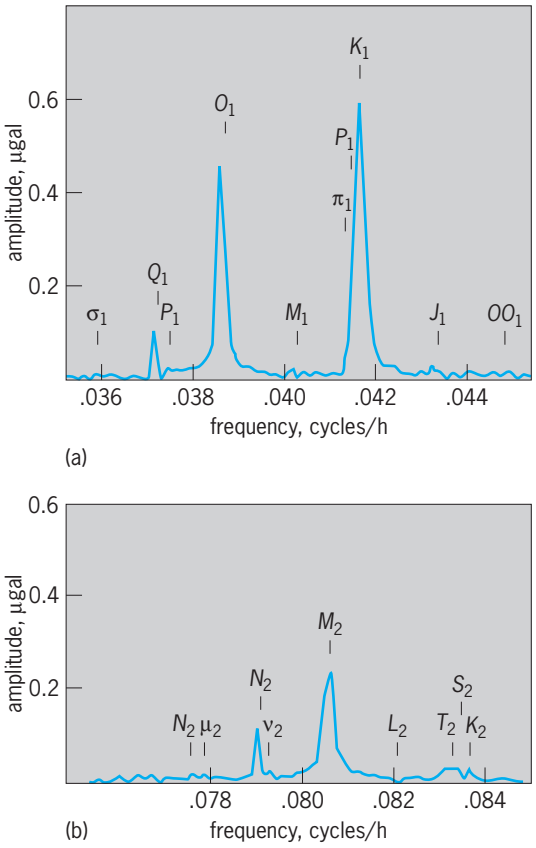


Fig. 3. Observed amplitude spectrum of (*a*) diurnal and (*b*) semidiurnal constituents at South Pole. 1 μgal = 1 μm/s^2. (*After B. V. Jackson and L. B. Slichter, The residual daily earth tides at the South Pole, J. Geophys. Res., 79(11):1711–1715, 1974*)

the tilt and gravimetric factors obtained by a number of workers at the nearly diurnal resonance of the solar waves ψ_1 and K_1 (**Fig. 3**).

By stacking qravity measurements, the complex eigenfrequency with strength, period, and damping of the nearly diurnal free wobble are determined. The estimate of the real part of the strength amplitude is in close agreement with the first experimental determination of the inner-pressure gravimetric factor.

Effects of inertia, ellipticity, and anisotropy. In an elliptical and rotating Earth, there is a coupling between the spherical and toroidal displacement fields in the core. In 1981 the Dahlen-Smith formulation of free oscillation of a rotating, elliptical Earth was adopted to calculate the gravimetric factors for M_2 and O_1 as a function of latitude, which reach a maximum at the Equator and a minimum at the poles. A uniform means of instrumental calibration may bring the observed values of later workers close to that given by the former method.

One intriguing problem in earth tidal studies is the effect of anisotropy in the Earth's mantle on the earth tide and ocean-load tide. Equations of deformation of the Earth with anisotropic regions through lagrangian mechanics are derived. The solution indicates that earth tides are virtually blind, at least to

the anisotropy of the upper mantle. However, load tides are affected by as much as 2.5%.

Earth tides and earthquakes. Volcanic earthquake swarms at Pavlof Volcano (southwestern Alaska Peninsula) correlated significantly with solid earth tidal stress rate for periods just before and just after explosive eruptions. In the 1974 minor eruption sequence, preeruptive earthquake swarms occurred during increasing earth tidal extension. Earth tides may have an effect on certain aspects of volcanic activity at Pavlof, and the polarity of the observed tidal correlation varies systematically during a volcanic eruption. *See* EARTHQUAKE.

Louis B. Slichter; John T. Kuo

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Earthquake

The sudden movement of the Earth caused by the abrupt release of accumulated strain along a fault in the interior. The released energy passes through the Earth as seismic waves (low-frequency sound waves), which cause the shaking. Seismic waves continue to travel through the Earth after the fault motion has stopped. Recordings of earthquakes, called seismograms, illustrate that such motion is recorded all over the Earth for hours, and even days, after an earthquake.

Characteristics

Earthquakes vary immensely in size, from tiny events that can be detected only with the most sensitive seismographs, to great earthquakes that can cause extensive damage over widespread areas. Although thousands of earthquakes occur every day, and have for billions of years, a truly great earthquake occurs somewhere in the world only once every year. When a great earthquake occurs near a highly populated region, tremendous destruction can occur within a few seconds. In 1976, 600,000 people were killed in Tangshan, China, by a single earthquake. The city of Lisbon, one of the principal capitals of that day, was utterly destroyed, with high loss of life, in 1755. In the twentieth century such cities as Tokyo and San Francisco have been leveled by earthquakes. In these more modern cases, much of the damage was not due to the shaking of the earthquake itself, but was caused by fires originating in the gas and electrical lines which interweave modern cities, and by damage to fire-fighting capability which rendered the cities helpless to fight the conflagrations.

Cause. The locations of earthquakes which occurred between 1957 and 2000 are shown on the map in **Fig. 1**. The map shows that earthquakes are not distributed randomly over the globe but tend to occur in narrow, continuous belts of activity. Approximately 90% of all earthquakes occur in these belts, which define the boundaries of the Earth's plates. The plates are in continuous motion with respect to one another at rates on the order of centimeters per year; this plate motion is responsible for most geological activity.

Plate motion occurs because the outer cold, hard skin of the Earth, the lithosphere, overlies a hotter, soft layer known as the asthenosphere. Heat from decay of radioactive minerals in the Earth's interior sets the asthenosphere into thermal convection. This convection has broken the lithosphere into plates which move about in response to the convective motion in a manner shown schematically in **Fig. 2**. The plates move apart at oceanic ridges. Magma wells up in the void created by this motion and solidifies to form new sea floor. This process, in which new sea floor is continually created at oceanic ridges, is called sea-floor spreading. Since new lithosphere is continually being created at the oceanic ridges by sea-floor spreading, a like amount of lithosphere must be destroyed somewhere. This occurs at the oceanic trenches, where plates converge and the oceanic lithosphere is thrust back down into the asthenosphere and remelted. The melting of the lithosphere in this way supplies the magma for the volcanic arcs which occur behind the trenches. Where two continents collide, however, the greater bouyancy of the less dense continental material prevents the lithosphere from being underthrust, and the lithosphere buckles under the force of the collision, forming great mountain ranges such as the Alps and Himalayas. Where the relative motion of the plates is parallel to their common boundary, slip occurs along great faults which form that boundary, such as the San Andreas fault in California. *See* ASTHENOSPHERE; EARTH, HEAT FLOW IN; LITHOSPHERE; MAGMA; VOLCANOLOGY.

According to the theory of plate tectonics, the motion of the plates is very similar to the movement of ice floes in arctic waters. Where floes diverge, leads form and water wells up, freezing to the floes and producing new floe ice. The formation of pressure ridges where floes converge is analogous to the development of mountain ranges where plates converge. *See* OROGENY; PLATE TECTONICS.

Stick-slip friction and elastic rebound. As the plates move past each other, little of the motion at their boundaries occurs by continuous slippage; most of the motion occurs in a series of rapid jerks. Each jerk is an earthquake. This happens because, under

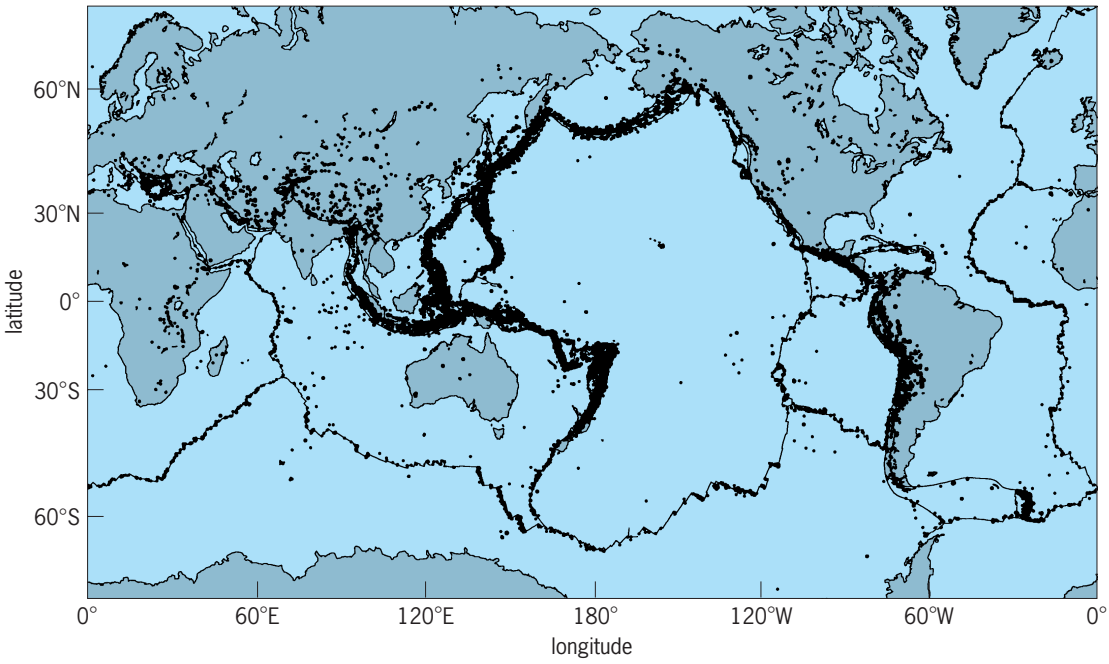


Fig. 1. Seismicity of the Earth from 1957 to 2000; depths to 700 km (435 mi). Earthquakes are plotted as circles; the plate boundaries are black. (*After E. R. Engdahl et al., Global teleseismic earthquake relocation with improved travel times and procedures for depth determination, Bull. Seis. Soc. Amer., 88:722–743, 1998; and A. Villaseñor et al., Toward a comprehensive catalog of global historical seismicity, Eos, 78:581, 583, 588, American Geophysical Union, 1997*)

the pressure and temperature conditions of the shallow part of the Earth's lithosphere, the frictional sliding of rock exhibits a property known as stick-slip, in which frictional sliding occurs in a series of jerky movements, interspersed with periods of no motion—or sticking. In the geologic time frame, then, the lithospheric plates chatter at their boundaries, and at any one place the time between chatters may be hundreds of years.

The periods between major earthquakes is thus one during which strain slowly builds up near the plate boundary in response to the continuous movement of the plates. The strain is ultimately released by an earthquake when the frictional strength of the plate boundary is exceeded. This pattern of strain buildup and release was discovered by H. F. Reid in his study of the 1906 San Francisco earthquake. During that earthquake, a 250-mi-long (400-km) portion of the San Andreas fault, from Cape Mendocino to the town of Gilroy, south of San Francisco, slipped an average of 12 ft (3.6 m). Subsequently, the triangulation network in the San Francisco Bay area was resurveyed; it was found that the west side of the fault had moved northward with respect to the east side, but that these motions died out at distances of 20 mi (32 km) or more from the fault. Reid had noticed, however, that measurements made about 40 years prior to the 1906 earthquake had shown that points far to the west of the fault were moving northward at a slow rate. From these clues, he deduced his theory of elastic rebound, illustrated schematically in **Fig. 3**. The figure is a map view, the vertical line representing the fault separating two moving plates. The unstrained rocks in Fig. 3*a* are distorted by the slow movement of the plates in Fig. 3*b*. Slippage in an earthquake, returning the rocks to an unstrained state, occurs as in Fig. 3*c*. *See* FAULT AND FAULT STRUCTURES.

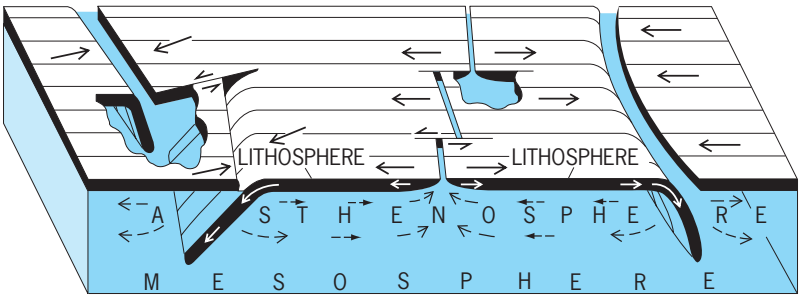


Fig. 2. Movement of the lithosphere over the more fluid asthenosphere. In the center, the lithosphere spreads away from the oceanic ridges. At the edges of the diagram, it descends again into the asthenosphere at the trenches. (*After B. Isacks, J. Oliver, and L. R. Sykes, Seismology and the new global tectonics, J. Geophys. Res., 73:5855–5899, American Geophysical Union, 1968*)

Classification. Most great earthquakes occur on the boundaries between lithospheric plates and arise directly from the motions between the plates. Although these may be called plate boundary earthquakes, there are many earthquakes, sometimes of substantial size, that cannot be related so simply to the movements of the plates.

Near many plate boundaries, earthquakes are not restricted to the plate boundary itself, but occur over a broad zone—often several hundred miles wide—adjacent to the plate boundary. These earthquakes, which may be called plate boundary–related earthquakes, do not reflect the plate motions directly, but are secondarily caused by the stresses set up at the plate boundary. In Japan, for example, the plate boundaries are in the deep ocean trenches offshore

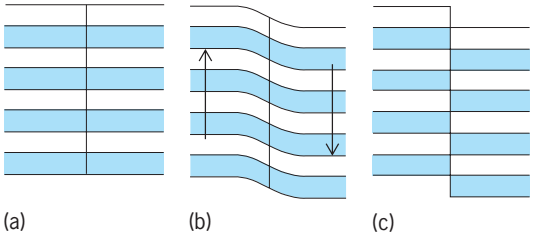


Fig. 3. Schematic of elastic rebound theory. (*a*) Unstrained rocks (*b*) are distorted by relative movement between the two plates, causing strains within the fault zone that finally become so great that (*c*) the rocks break and rebound to a new unstrained position. (*After C. R. Allen, The San Andreas Fault, Eng. Sci. Mag., Calif. Inst. Technol., pp. 1–5, May 1957*)

of the Japanese islands, and that is where the great plate boundary earthquakes occur. Many smaller events occur scattered throughout the Japanese islands, caused by the overall compression of the whole region. Although these small events are energetically minor when compared to the great offshore earthquakes, they are often more destructive, owing to their greater proximity to population centers.

Although most earthquakes occur on or near plate boundaries, some also occur, although infrequently, within plates. These earthquakes, which are not related to plate boundaries, are called intraplate earthquakes, and can sometimes be quite destructive. The immediate cause of intraplate earthquakes is not understood. Some of them can be quite large. Three of the largest earthquakes known to have occurred in the United States were part of a sequence of intraplate earthquakes which took place in the Mississippi Valley, near New Madrid, Missouri, in 1811 and 1812. Another intraplate earthquake, in 1886, caused moderate damage to Charleston, South Carolina.

In addition to the tectonic types of earthquakes described above, some earthquakes are directly associated with volcanic activity. These volcanic earthquakes result from the motion of undergound magma that leads to volcanic eruptions.

Sequences. Earthquakes often occur in well-defined sequences in time. Tectonic earthquakes are often preceded, by a few days to weeks, by several smaller shocks (foreshocks), and are nearly always followed by large numbers of aftershocks. Foreshocks and aftershocks are usually much smaller than the main shock. Volcanic earthquakes often occur in flurries of activity, with no discernible main shock. This type of sequence is called a swarm.

Size. Earthquakes range enormously in size, from tremors in which slippage of a few tenths of an inch occurs on a few feet of fault, to the greatest events, which may involve a rupture many hundreds of miles long, with tens of feet of slip. Accelerations exceeding 1 *g* (acceleration due to gravity) can occur during an earthquake. The velocity at which the two sides of the fault move during an earthquake is only 1–10 mi/h (10–100 cm/s), but the rupture front spreads along the fault at a velocity of nearly 5000 mi/h (3 km/s). The earthquake's primary damage is due to the generated seismic waves, or sound waves which travel through the Earth, excited by the rapid movement of the earthquake. The energy radiated as seismic waves during a large earthquake can be as great as 10^{12} cal (4.19×10^{12} joules), and the power emitted during the few hundred seconds of movement as great as a billion megawatts.

The size of an earthquake is given by its moment: average slip times the fault area that slipped times the elastic constant of the Earth. The units of seismic moment are dyne-centimeters. An older measure of earthquake size is magnitude, which is proportional to the logarithm of moment. Magnitude 2.0 is about the smallest tremor that can be felt. Most destructive earthquakes are greater than magnitude 6; the largest shock known was the 1960 Chile earthquake, with a moment of 10^{30} dyne-centimeters (10^{23} newton-meters) or magnitude 9.5. It involved a fault 600 mi (1000 km) long slipping 30 ft (10 m). The magnitude scale is logarithmic, so that a magnitude 7 shock is about 30 times more energetic than one of magnitude 6, and 30×30, or 900 times, more energetic than one of magnitude 5. Because of this great increase in size with magnitude, only the largest events (greater than magnitude 8) significantly contribute to plate movements. The smaller events occur much more often but are almost incidental to the process.

The intensity of an earthquake is a measure of the severity of shaking and its attendant damage at a point on the surface of the Earth. The same earthquake may therefore have different intensities at different places. The intensity usually decreases away from the epicenter (the point on the surface directly above the onset of the earthquake), but its value depends on many factors and generally increases with moment. Intensity is usually higher in areas with thick alluvial cover or landfill than in areas of shallow soil or bare rock. Poor building construction leads to high intensity ratings because the damage to structures is high. Intensity is therefore more a measure of the earthquake's effect on humans than an innate property of the earthquake.

Effects. Many different effects are produced by earthquake shaking. Although the fault motion that produced the earthquake is sometimes observed at the surface, often other earth movements, such as landslides, are triggered by earthquakes. On rare occasions the ground has been observed to undulate in a wavelike manner, and cracks and fissures often form in soil. The flow of springs and rivers may be altered, and the compression of aquifers sometimes causes water to spout from the ground in fountains. Undersea earthquakes often generate very long-wavelength water waves, which are sometimes called tidal waves but are more properly called seismic sea waves, or tsunami. These waves, almost imperceptible in the open ocean, increase in height as they approach a coast and often inflict great damage to coastal cities and ports. *See* LANDSLIDE; TSUNAMI.

Prediction

Earthquake prediction research has been going on for nearly a century. A successful prediction, specifying the time, location, and magnitude of an

earthquake, would save lives and billions of dollars in housing and infrastructure costs. Unfortunately, successful earthquake predictions are extremely rare. There are two basic categories of earthquake predictions: forecasts (months to years in advance) and short-term predictions (hours or days in advance). Forecasts are based a variety of research, including the history of earthquakes in a specific region, the identification of fault characteristics (including length, depth, and segmentation), and the identification of strain accumulation. Data from these studies are used to provide rough estimates of earthquake sizes and recurrence intervals.

An example of an earthquake forecast is the identification of seismic gaps, portions of the plate boundaries that have not ruptured in a major earthquake for a long time. These regions are most likely to experience great earthquakes in the future. **Figure 4** shows seismic gaps for the circum-Pacific region and indicates which gaps were most likely to experience a large or great earthquake in 1989–1999. Large earthquakes did occur in several of the likeliest gaps, but many large earthquakes occurred in less likely gaps as well. Earthquake probability estimates are another example of forecasting. Geologic, geodetic, and seismic information are combined to estimate the frequencies of damaging earthquakes in a specific region. Recent regional earthquake probability estimate studies have resulted in forecasts of an 80–90% probability of a magnitude 7 or larger earthquake in the southern California region before 2024, and a 70% probability of a magnitude 6.7 or larger earthquake in the San Francisco Bay region before 2030.

Short-term earthquake prediction is still entirely in the realm of ongoing research, and no method is known to be reliable. Evidence is emerging that short-term prediction may be inherently impossible due to the complex and chaotic nature of the earthquake process.

Christopher H. Scholz; Kaye M. Shedlock

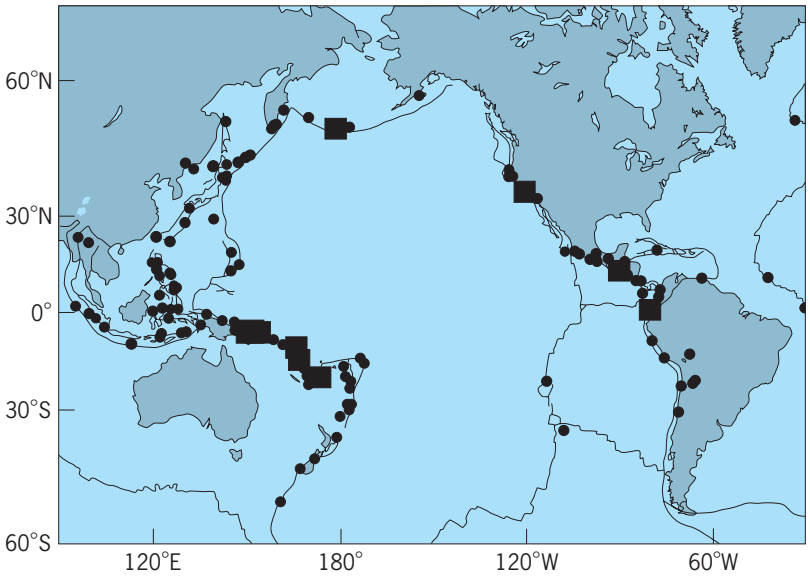


Fig. 4. Major seismic gaps, western Pacific. The plate boundaries are shown in black. Black squares mark the locations of seismic gaps with a 50% or greater chance of being filled by a large earthquake between 1989 and 1999. Black dots are magnitude 7.0 or larger earthquakes that occurred between 1989 and 1999. (*After S. P. Nishenko, Pure Appl. Geophys., 135:169–259, 1991*)

Deep Earthquakes

Most earthquakes occur at depths shallower than about 50 km (30 mi) [**Fig. 5***a*] and are usually found near plate boundaries. A few percent of all shocks occur at depths of 300–700 km (183–427 mi), depths that correspond to earth pressures of 100,000–250,00 atm (1–2.5 × 10^{10} Pa; Fig. 5*b* and **Fig. 6**). That the mantle can suddenly rupture rather than flow plastically at such conditions has elicited wonder since deep earthquakes were first discovered in the 1920s. Modern insight into these phenomena has come from scientific advances of plate tectonics, seismic tomography, and the mineral physics of the deep mantle based on very high pressure experiments on mantle minerals.

Most, if not all, deep earthquakes occur in inclined belts inside slabs, the cold, dense, and strong lithospheric plates that dive deeply into the Earth's mantle in places where plates are converging. Seismic waves have been used to image variations in the seismic wave speeds in the Earth. These anomalies in seismic tomographic images reflect differences in temperature, mineralogy, or composition. As expected, deep earthquakes occur in the high-wave-speed anomalies that mark cold slabs, anomalies that have been traced to depths as great as 2000 km (1220 mi) or more.

Curiously, earthquakes occur no deeper than 650–700 km (397–427 mi), far shallower than the maximum depths to which slabs descend. This abrupt shutoff and the gradual onset of the deep earthquake population at 300–350 km (183–214 mi) bracket approximately the transition zone of the mantle where seismic wave speeds abruptly increase (Fig. 6*a*). High-pressure experiments indicate that the mineralogy of the mantle changes at those depths and pressures from upper-mantle mineralogy (dominantly olivine and pyroxenes) to the minerals spinel, ilmenite, and majorite in the transition zone and, in turn, to the lower-mantle perovskite and oxide minerals. Slab mantle penetrating through the transition zone is expected to transform to these denser minerals.

Most deep earthquakes occur in the depth interval of the transition zone where upper-mantle slab minerals are reconstructed to their denser structural forms. Attention has therefore been drawn to the possibility that deep earthquakes are somehow caused by the mineralogical transformation of slabs as they descend into and through this region. Early speculation was that deep earthquakes represent rapid implosions that might occur when slab minerals transform suddenly to their denser, high-pressure forms. The patterns of seismic waves that radiate from deep earthquake sources indicate, however, that such disturbances represent slip on a fault, as do shallow earthquakes. If a connection exists between deep earthquakes and mantle phase changes,

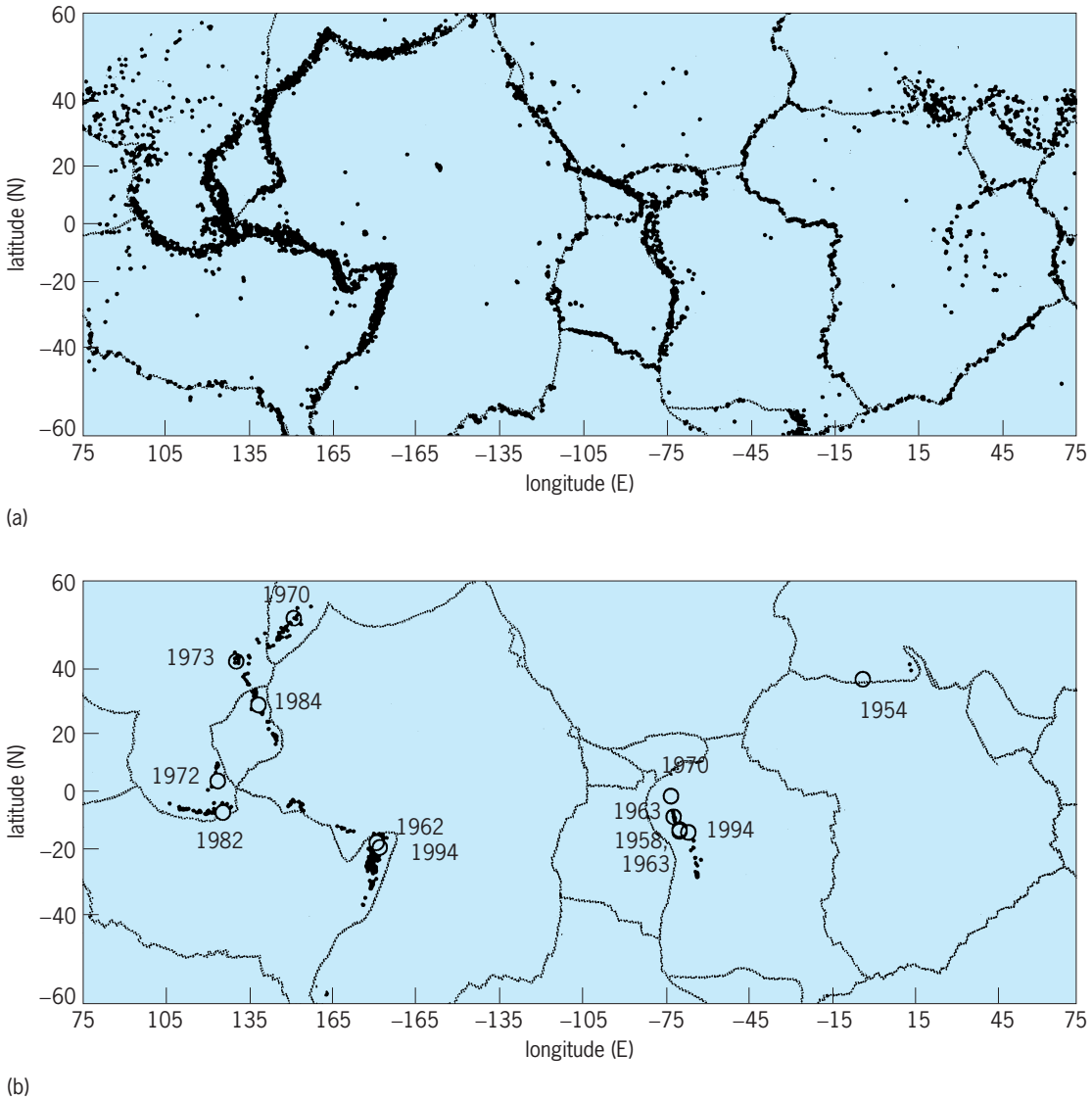
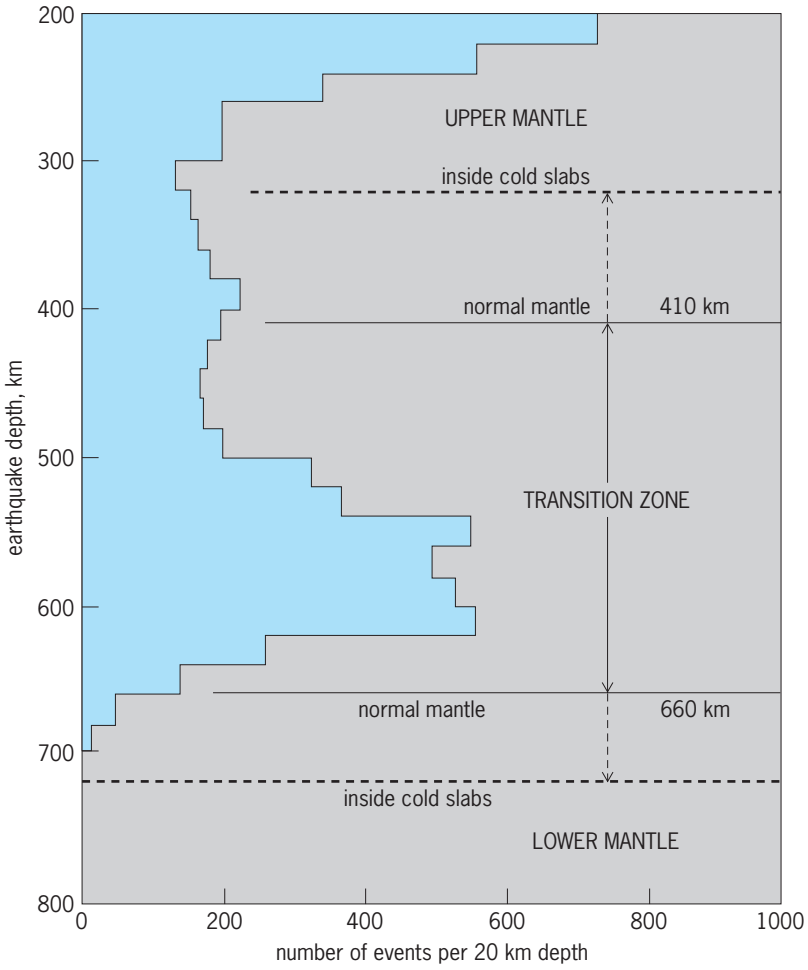


Fig. 5. Map showing the global distribution of shallow and deep earthquakes. (*a*) Earthquakes shallower than 40 km (25 mi). (*b*) Earthquakes deeper than 300 km (186 mi; open circles with dates indicate the locations of the 13 largest recorded deep shocks, including the 1994 Bolivian event).

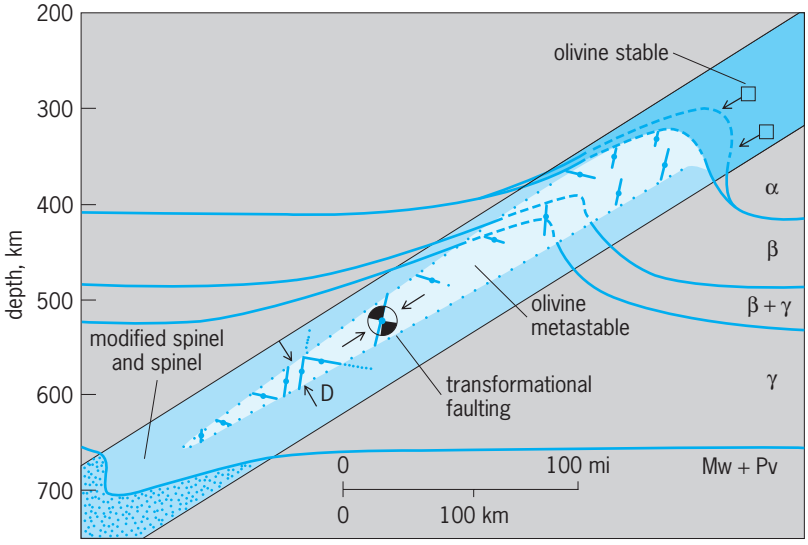
the underlying process must facilitate failure by faulting.

A clue to the nature of this possible connection comes from the observation that deep earthquakes do not occur in all slabs, only in those that are very cold because they are descending at very fast rates. Low slab temperatures are important because such conditions favor the metastable persistence of upper-mantle minerals in the coldest interiors of slabs as they descend into the transition zone (Fig. 6*b*). Laboratory studies show that some minerals deformed under metastable conditions will rupture by an unusual shear instability in which the mineral is transformed to denser minerals in the shear zone. This shear instability, called transformational faulting, is not inhibited by high pressures and hence is an attractive candidate for the faulting mechanism of deep earthquakes. According to this theory, deep earthquakes do not occur in the lower mantle because low-density upper-mantle slab rocks are too buoyant to sink into the lower mantle.

An extraordinary demonstration of the potential scale of deep seismic faulting was demonstrated by the great deep earthquake that occurred at a depth of about 640 km (390 mi) beneath the Amazonian rainforest of Bolivia on June 9, 1994. The ground motions of this magnitude 8.2 shock, the largest deep event on record, were felt by people in places as far north as Toronto and Seattle, corresponding to surface distances from the earthquake focus exceeding 8000 km (4880 mi). Theoretical analysis of the seismograms written by this earthquake indicate that deep rupture occurred on a horizontal fault over an area of about 1600–2500 km^2 (618–965 mi^2) with a rupture duration of about 40–50 s and a shear offset of about 10–15 m (33–49 ft). This great, deep rupture occurred in the Nazca plate subducting beneath the South American plate, and was so energetic that it excited Earth's free oscillation modes and the planet rang like a low-frequency bell for many weeks after



(a)



(b)

Fig. 6. Depth histogram of earthquakes compared to the mineralogical structure of the mantle. (*a*) Depth histogram of well-located earthquakes (1964–1991) in relationship to seismic-wave speed discontinuities caused by mineralogical changes in normal mantle and in cold slabs. (*b*) Hypothetical mineralogical structure of very cold slabs and normal mantle, emphasizing the phase changes associated with the olivine component [$(Mg,Fe)_2SiO_4$] of the mantle [α (olivine), β (modified spinel) and γ (spinel)]; Mw + Pv are magnesiowüstite and perovskite, the higher-pressure minerals dominating the lower mantle. Transformational faulting is a shear instability that can occur in metastable olivine under stress and can produce deep earthquakes.

the earthquake. The global occurrence of such earthquakes is the deepest direct expression of Earth's thermal convection system. *See* EARTH, CONVECTION IN.

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Earthquake engineering

The branch of engineering concerned with reducing earthquake or seismic risk to structures. Because strong earthquakes are rare events, building codes have traditionally allowed a significant degree of damage. Even high-seismic regions, such as San Francisco or Tokyo, typically experience a strong earthquake only once in many decades. If an earthquake occurs, buildings and other structures are designed such that most will be damaged (but should not collapse), and will have costs for repairs, business interruption, and potentially casualties. While building collapse is the primary cause of loss of life in most earthquakes, other contributors to earthquake loss include equipment and contents damage, business interruption, and damage to lifelines, such as water, power, gas, communications, and transportation. To limit these losses to acceptable levels, earthquake engineering involves a process of (1) seismic hazard identification, (2) structural analysis, design, and/or retrofitting to prevent structural collapse and reduce property damage, and (3) review of equipment and operations to prevent disruption due to earthquakes—that is, an integrated, comprehensive program of facility seismic review, analysis, retrofit, emergency planning, and risk transfer, involving the expertise of mechanical engineers, operations specialists, emergency planners, and insurers, in addition to geoscientists and structural engineers.

Hazards. Most earthquakes are caused by the fracture and sliding of portions of the Earth's crust along faults, which may be hundreds of miles long, from 1 mi (1.6 km) to over 100 mi (160 km) deep, and

sometimes not readily apparent on the ground surface. Earthquakes can occur anywhere on Earth, but most occur along major tectonic plate boundaries ("Ring of Fire"), especially on the circum-Pacific plate boundary, the Caribbean, and the Trans-Alpide belt (stretching from southern France through the Mediterranean and the Middle East, along the Himalayan foothills and the Indonesian archipelago). In the United States, 39 of the 50 states are considered at moderate to high seismic risk, with major earthquake potential in the western states, Alaska, the central United States (St. Louis–Memphis region), and portions of the east coast (South Carolina, Massachusetts). *See* EARTHQUAKE; SEISMIC RISK.

Earthquakes can cause significant damage to the built environment due to fault rupture, vibratory ground motion (shaking), inundation (tsunami, seiche, dam failure), various kinds of permanent ground failure (liquefaction, landslide), and fire or hazardous materials release. In a particular event, any of these hazards can dominate, and historically each has caused major damage and great loss of life. The expected damage given a specified value of a hazard parameter is called vulnerability, and the product of the hazard and the vulnerability is the expected loss or seismic risk.

For most earthquakes, shaking is the dominant and most widespread cause of damage. Shaking near the actual earthquake rupture lasts only during the time when the fault ruptures, a process which takes seconds or at most a few minutes. The seismic waves generated by the rupture propagate long after the movement on the fault has stopped, spanning the globe in about 20 min. Typically earthquake ground motions are powerful enough to cause damage only in the near field (that is, within a few tens of kilometers from the causative fault). In a few instances, long period motions have caused significant damage at great distances to selected structures. A prime example was the 1985 Mexico City earthquake, where a magnitude 8.1 earthquake occurring at a distance of approximately 400 km from Mexico City caused numerous collapses of mid- and high-rise buildings.

The most important scales for measuring earthquake magnitude are surface-wave magnitude M_s, body-wave magnitude m_b, and moment magnitude M_W (the last is the more commonly preferred scale today). Magnitude can be related to the total energy release by $\log_{10} E = 11.8 + 1.5M_S$, where E is the total energy in ergs. Since $10^{1.5} = 31.6$, an increase of one magnitude unit is equivalent to 31.6 times more total energy release, an increase of two magnitude units ~1000 times more total energy, and so forth. Whereas magnitude is a measure of the overall size of a single earthquake, intensity is a measure of the effect, or the strength, of an earthquake hazard at a specific location. Intensity scales in use include the Modified Mercalli Intensity (MMI) in the United States, Medvedev-Sponheur-Karnik (MSK-81) in Europe, and the Japan Meteorological Agency (JMA) scale. Roman numerals are traditionally used for intensity scales to indicate the qualitative nature of the scales, which are based on subjective observations rather than instrumental records. For the MMI and MSK scales, 0 is no earthquake, VI is the initiation of damage to poor-to-average structures, and XII is total destruction. Earthquake shaking at a site is recorded on a seismometer, which digitally or optically records a time history of ground accelerations at the site.

Fig. 1. Relatively new reinforced concrete buildings that collapsed in the Marmara, Turkey, earthquake, August 17, 1999.

Statistical analysis of hundreds of such accelerograms provide attenuation regressions, which permit estimation of the maximum acceleration, velocity, or displacement that a structure will experience at a site as a function of the structure's natural period, the site-specific soil properties, and the magnitude, distance, and depth of the hypothesized earthquake. In the United States, ground motions due to all potential earthquakes have been synthesized by the U.S. Geological Service, so that any site's seismic hazard (based on an average soil type) can be quickly determined. Based on this hazard, or on building code requirements mapped at a regional scale, the seismic lateral force requirements against which a structure must be designed are determined. *See* SEISMOLOGY.

Structural analysis and design. In the developed world, existing buildings and infrastructure constitute by far the preponderance of earthquake risk, so that a major focus of earthquake engineering is on identification, analysis, and mitigation (reduction) of this risk. New construction is generally safer, but is still a focus for earthquake engineers, especially larger or unusual structures. In the developing world, even recent and new construction is a significant contributor to seismic risk, due to lack of building code enforcement (**Fig. 1**). The most seismically hazardous existing building structures are low-strength masonry (such as adobe in developing countries), unreinforced brick masonry (such as exists in older portions of United States cities), nonductile reinforced concrete (typically, concrete frames constructed prior to the 1980s), and certain kinds of precast concrete buildings. Steel and wood structures are not immune to earthquake damage, but typically are less collapse-prone. However, the tallest high-rise building to collapse, for any reason, was the 23-story Piño Suarez steel-framed building in the 1985 Mexico City earthquake. Building vulnerability is typically mitigated via structural retrofits involving strengthening wall-diaphragm connections, adding shear walls or bracing, and improving ductility of columns by steel jacketing or other methods. A diaphragm is a horizontal force-distributing element, such as a roof or floor.

Earthquake analysis and design of a structure is commonly performed assuming the structure remains elastic (that is, the design neglects the effects of damage), though inelastic analyses (which takes into account material inelasticity) are increasingly used. For ordinary design, the actual earthquake forces are reduced by a response modification factor that varies depending on the assumed inherent ductility of the structure, based on its lateral force-resisting system (moment frame, braced, shear wall) and material (wood, steel, concrete). This force reduction, along with appropriate safety factors, is combined with gravity and other loads in the design of the overall structure. Pseudo-static or linear dynamic analytical methods are used to determine structural member forces. Pseudo-static methods are based on the structure's natural period (first mode of vibration) and are appropriate only for simple structures. Linear dynamic methods account for the first and higher modes, which analyses are usually more quickly performed in the frequency domain, using techniques based on the fast Fourier transform. For larger and more important structures, nonlinear dynamic analyses are used, typically in the time domain. In a nonlinear dynamic analysis, the structure is subjected to earthquake acceleration time histories (actual or synthesized records, scaled to match the site hazard), and member response into the inelastic range is taken into account, including P-Δ effects (that is, the increase in overturning moment due to the structure's weight P times its lateral deflection Δ). Dynamic analysis has become common due to the advent of more powerful computers and specialized software—ETABS, SAP2000, ANSYS, STAAD, and LARSA are some of the structural analysis packages commonly used today. *See* BUILDINGS; LOADS, DYNAMIC; STRUCTURAL ANALYSIS.

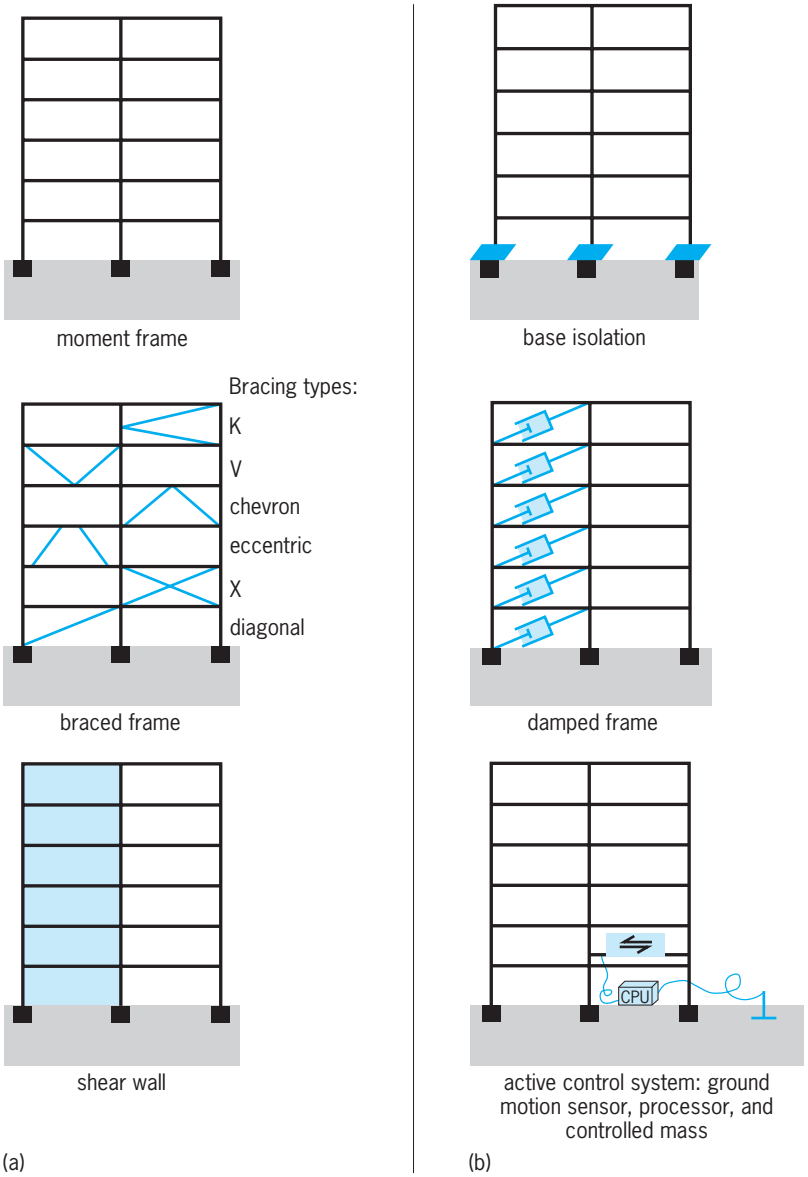

Fig. 2. Structural designs: (*a*) Traditional earthquake force-resisting systems. (*b*) Emerging technologies for earthquake force-reducing systems.

New design approaches. Traditionally, engineers have designed structures to resist earthquake forces, intending the strength of structural members to be equal to or greater than the seismic demand placed on them (**Fig. 2**). During the 1980s and 1990s, new approaches to seismic design emerged, which involved modifying the structural response to reduce earthquake loads to more tolerable levels. These included base isolation, supplemental damping, and active control. Base isolation involves placing special components, termed isolators (not always at the base, so that the technique is more properly termed structural isolation), within the structure that are relatively flexible in the lateral direction, yet can sustain the vertical load. When the earthquake causes ground motions beneath the structure, the isolators allow the structure to respond more slowly than it would without them, resulting in lower seismic demand on the structure. Isolators may be laminated steel with high-quality rubber pads, sometimes incorporating lead or other energy-absorbing materials, or parabolic dish-shaped base plates, which rely on the structure's own weight trying to "climb" the sloping sides of the "dish" to counteract the lateral force of the earthquake. Supplemental damping involves placing dampers within the structure, which retard the structural response of the normal or lightly damped structure, again resulting in lower seismic demand on the structure. Active control involves placing hydraulic rams or other actuating devices within the structure, which introduce forces counter to those caused by the earthquake, thus negating some or all of the seismic demand. To determine how much force to apply to an actuator involves a real-time solution of the dynamics of the structure. Sensors must measure the ground motion at the base of the structure, a structural analysis for these accelerations must be performed on a dedicated processor, and the optimum pattern of forces must be determined and applied via quick-response actuators during a time interval equal to or less than the dynamic response of the structure. Base isolation and supplemental damping are currently entering the mainstream of structural engineering, while active control is still largely the subject of research.

The most recent development in earthquake engineering is performance-based design, in which the expected structural damage due to the maximum expected earthquake is quantified. If the expected damage is unacceptably high, the owner and the engineer may agree on a design in excess of building code requirements. *See* ARCHITECTURAL ENGINEERING; CIVIL ENGINEERING.

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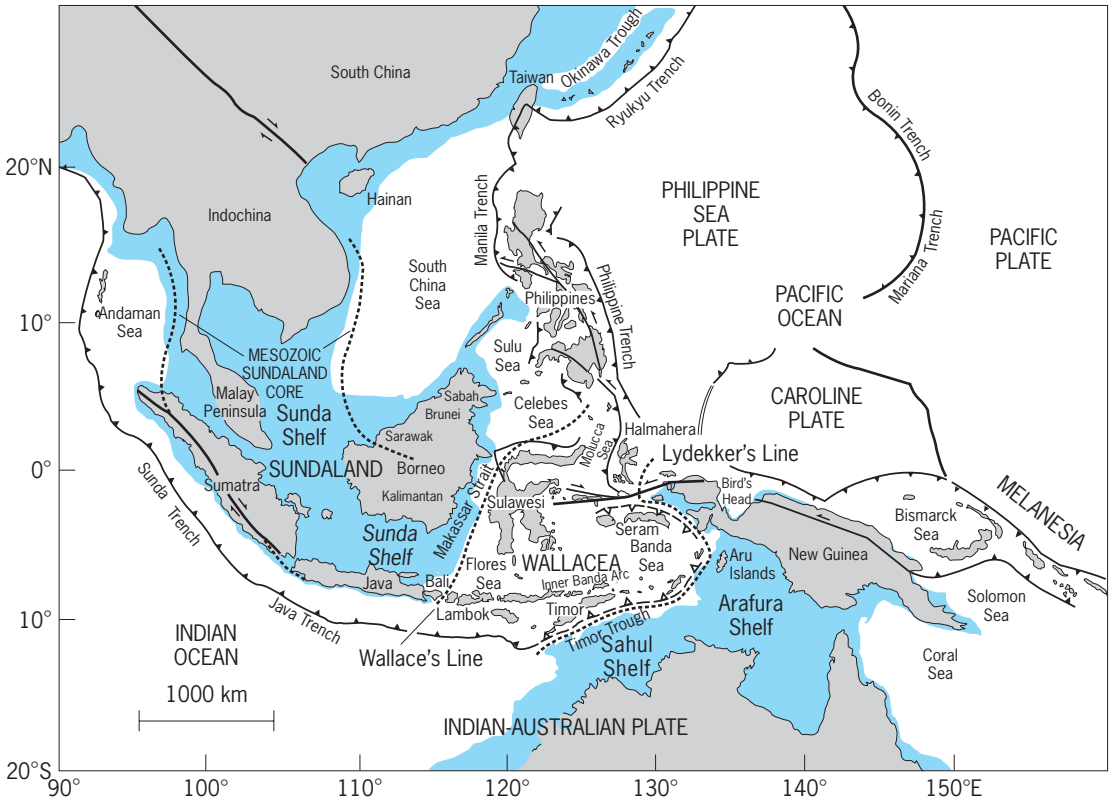
East Indies

A loosely defined region in southeast Asia broadly comprising former colonial possessions of European nations and now forming the countries of Malaysia, Brunei, and Indonesia. The islands of the East Indies extend for about 2800 mi (4500 km) from western Sumatra to New Guinea. They form part of a region of great geological and biological diversity.

Geography. Three broad geographic subdivisions can be made (see **illus.**): (1) Sundaland, which comprises the shallow marine Sunda Shelf, peninsular southeast Asia, and the islands west of the Makassar and Lombok straits; (2) northern Australia, which includes the shallow marine Sahul and Arafura shelf areas and islands surrounding northern Australia and New Guinea; and (3) Wallacea, which comprises the islands south of the Philippines between Sundaland and northern Australia. Each division is an important geomorphological and biological region. The boundaries of Sundaland and northern Australia can be drawn at the 660-ft (200-m) bathymetric contour. Deep trenches with depths averaging 5 mi (8 km) border the margins of Sundaland with the Indian Ocean, and northern Australia with the Pacific Ocean. In contrast, within the area of Wallacea there are a number of deep troughs, and there are large variations in relief with volcanic and nonvolcanic mountains, typically up to 6600–9900 ft (2000–3000 m) in height, separated by deep marine basins underlain by extended continental and oceanic crust with depths of several kilometers. *See* ASIA; AUSTRALIA.

Landscape. Sundaland is very flat with a mean elevation only slightly above sea level. A shallow shelf currently lies beneath the sea between Borneo, Java, Sumatra, and the Malay Peninsula, and has been intermittently emergent in the past, notably during Pleistocene glaciations. Highlands are found in the Malay Peninsula, Sumatra, Java, and Borneo (comprising Brunei, the Malaysian states of Sabah and Sarawak, and Indonesian Kalimantan). The central part of the Malay Peninsula has a western range of granite highlands up to 7260 ft (2200 m) which provided the once important tin deposits of Malaysia. Further east are lower peaks, again mainly underlain by intrusive granitic rocks. In Sumatra, the narrow Barisan mountain range runs the length of the island on its western side, within which is the major active Sumatran strike-slip fault zone, and on which is built a chain of volcanoes with heights up to 12,550 ft (3800 m). Most of eastern Sumatra is swampy alluvial lowland. Java is one of the richest agricultural islands of the tropics. There are more than 40 young volcanoes with heights up to almost 12,200 ft (3700 m) forming a continuous spine running the length of the island. In contrast to Sumatra and Java, Borneo has a broad low central mountain range running from southwest to northeast which becomes higher toward the northeast. The border of Kalimantan with Sarawak and Sabah essentially follows the drainage divide, and there are a dozen or so peaks above 6600 ft (2000 m), of which the highest is Mount Kinabalu



East Indies. The colored areas represent the shallow continental shelves of Sundaland and northern Australia, drawn at the 660-ft (200-m) isobath. The western boundary of Wallacea is Wallace's Line of 1863. The eastern boundary is Lydekker's Line, which is the western limit of strictly Australian fauna. Major tectonic plates are named. The Asian plate is a composite of several smaller plates.

in Sabah (13,525 ft or 4098 m). Kinabalu is formed of young granite, testifying to very rapid and recent uplift, and its glacial landscape at the summit shows that it was the site of a Pleistocene icecap. Nearly half of Borneo is lowlands under 660 ft (200 m), which are frequently swampy. In Borneo the largest rivers of Sundaland drain west and north into the Makassar Strait, and the South China, Sulu, and Celebes seas, where they deposited immense volumes of sediment in late Cenozoic deltas which are very rich in oil and gas.

Like Sundaland, northern Australia has a mean elevation close to sea level except in New Guinea, the world's second largest island, which is nearly 1350 mi (2200 km) long. The high mountains of the central part of the island are a Tertiary fold and thrust belt rising to more than 16,500 ft (5000 m). Many of them are snow-capped, even though they are within 5° of the Equator, and display the effects of Pleistocene glaciation. North of the central mountains there is a broad valley with a lower coastal range to its north. The rapid uplift of the mountains and the heavy tropical rainfall mean that there are many fast-flowing rivers. The rivers have very high sediment yields, and those of New Guinea are estimated to carry as much sediment in total as all those draining North America.

Within Wallacea are numerous small rugged islands, all of which have risen above sea level in the last few million years. The largest is Sulawesi, in which most land is above 1485 ft (450 m) and where very young central mountains rise to over 11,000 ft (3400 m). Numerous lakes occupy depressions within young strike-slip fault zones; Lake Poso is 4950 ft (1500 m) deep. Timor and Seram, although smaller, are similarly mountainous, with many peaks above 5950 ft (1800 m) and maximum heights on both islands close to 9900 ft (3000 m).

Climate and ocean circulation. At present, almost all the islands lie in a belt close to the Equator within the Intertropical Convergence Zone (ITCZ). An equatorial climate prevails, with high rainfall and, except at higher elevations, high temperatures throughout the year. Diurnal variations are greater than the difference of mean temperatures of the hottest and coldest months. High relative humidity is normal in most lowland regions. Regionally, there are significant variations in rainfall, reflecting topography and position with respect to major landmasses and oceans, and each island's climate can be different. Borneo is the only large island within which there is broadly an ever-wet tropical climate. Elsewhere, monsoons, reflecting movement of air masses between Asia, Australia, and the surrounding oceans, control seasonal changes in winds and variations in precipitation. A tropical climate with typically more than 120 in. (3000 mm) of rainfall each year prevails in Sumatra, western Java, Borneo, Sulawesi, and the Moluccas; but extended dry seasons occur in eastern Java and the Lesser Sunda Islands. Climatic and

topographic variation is reflected in a considerable variety in vegetation, from tropical rainforest to savanna, now further altered by human activity. *See* CLIMATOLOGY; TROPICAL METEOROLOGY.

The shallow seas with narrow deep-water passages of Wallacea mean that the region is particularly important for oceanic circulation. The Indonesian throughflow is a current to the Indian Ocean which transports large amounts of warm water from the Pacific, influencing sea surface temperatures, salinity, and rainfall. The magnitude and variations of this current are important controls on the thermohaline balance of the Pacific and Indian oceans, and perhaps on global thermohaline circulation. Most water passes from the North Pacific via the Celebes Sea, Makassar Strait, Flores Sea, and Banda Sea. Less important volumes of water move from the South Pacific east of Halmahera into the Banda Sea. Changes in the relative importance of these currents are probably seasonal and in the longer term may vary with El Niño. The importance of the present throughflow suggests that the convergence of Australia and Asia in the last few million years may have influenced local and possibly global climate, as the Pacific-Indian seaway has become increasingly restricted. *See* EQUATORIAL CURRENTS; OCEAN CIRCULATION.

Geological history. The East Indies are characterized by intensely active seismicity and volcanic activity. The correlation of seismicity, volcanicity, deep trenches, and strong negative gravity anomalies along the Sunda and Banda arcs was noted long before the formulation of the theory of plate tectonics; these are now known to be features characteristic of the subduction of oceanic lithosphere. The history of convergence of the Pacific, Indian-Australian, and Asian plates offers a broad explanation of the geological development and complexity of the region, but many small plates also need to be considered. However, the full details of the development of the region are still far from understood because of its size, relative inaccessibility, and the nature of the terrain. *See* CONTINENTAL DRIFT; GEODYNAMICS; PLATE TECTONICS.

In general, the region owes its origin to the pre-Cenozoic breakup of the Gondwana supercontinent, the subsequent movement of Gondwana fragments northward, and their eventual collision with Asia. Many fragments separated from Gondwana and amalgamated in southeast Asia over a considerable period of time. The process of rifting led to formation of new oceans, and the northward motion of Gondwana fragments required subduction of older oceanic crust at the edges of the growing Asian continent. By the late Mesozoic, fragments derived from Gondwana formed a composite Sundaland core surrounded by subduction zones.

Two major fragments, India and Australia, separated from Gondwana in the Cretaceous and moved northward as parts of different plates. India initially collided with the Asian continent about 50 million years ago, but continued to move north accompanied by complex internal deformation within Indochina and mainland southeast Asia, which continues to the present day. To the east of India, long-term subduction of oceanic lithosphere of the Indian plate produced the arcuate system of volcanoes and seismicity which extends through Sumatra and Java. Australia moved more slowly northward and during the last 25 million years has made an indirect collision with a composite southeast Asia which includes some of the Gondwana fragments that arrived earlier, and also the island arcs formed due to the subduction of oceanic crust north of Australia. Further east, arc-continent collisions have resulted from elimination of small oceanic basins formed above subduction zones north of Australia; this system of arcs and oceanic basins can be traced east along the Australian margin into Melanesia. Fragments of these arcs have been dispersed in New Guinea, east Indonesia, and the Philippines by the movement of the Pacific plate.

In east Indonesia the northward movement of Australia during the Cenozoic has been marked by arc-continent collision, major strike-slip motion within the north Australian margin in northern New Guinea, and collision of continental fragments derived from Australia. The Banda arc exemplifies many of the complexities of the region. There is the unusual 180° curvature of the arc. Deformation is very young and still active, and high mountains on land are separated by deep ocean basins. On its north side, Australian fragments are moving westward by strike-slip faulting. On its south side, young mountains in Timor contain deep-sea Australian margin sediments uplifted to elevations of more than 6600 ft (2000 m) within the last 3 million years following arc-continent collision. At the center of the arc is the Banda Sea, which is a young oceanic basin produced by very rapid extension, probably driven by subduction. The Banda region offers a snapshot of the processes which have been active in the entire region for many tens of millions of years. As a result, the region is geologically very complex and includes fragments of Gondwana continental origin, volcanic island arcs, and oceanic crust, all of which have been deformed by collision with the Sundaland core and then by subsequent collisions which have continued to the present day. However, despite the long-term overall convergence of the Asian, Australian, and Pacific plates, there has been frequent opening of small oceanic basins as well as extension related to strike-slip movements at plate boundaries within the region; the Banda Sea is one such basin. *See* MARINE GEOLOGY.

Plants and animals. Climatic changes during the past 20 million years have significantly influenced the distribution of many plants and animals in the region. The northward motion of Australia caused northern Australia and New Guinea to enter the tropical belt during this period. Collisions in eastern Indonesia closed ocean seaways and altered circulation between the Pacific and Indian oceans. At the same time, atmospheric circulation was affected by the rise of the Himalayas and the later rise of mountains in Borneo, Sulawesi, and New Guinea. About 20 million years ago, tropical ever-wet climates became established in much of the region, followed by a change

to monsoonal climates in many areas about 5 million years ago. More recently, during the Pleistocene, sea levels fluctuated dramatically and rapidly as a result of glacial cycles, which also caused considerable variations in rainfall and temperature with consequent effects on plant and animal life. Shallow parts of the continental shelves were exposed when sea levels were low, and land bridges formed between islands. *See* PLEISTOCENE.

Today, the waters of southeast Asia contain the highest marine faunal diversity in the world, and the islands of the East Indies contain some of the most diverse collections of plant and animal species. Tropical forests support more species than any other terrestrial ecosystem, and southeast Asia therefore includes some of the most important biological resources of the planet as well as a complex biogeography reflecting the many factors that have influenced the present distributions of plants and animals. The East Indies have been a critical region in the development of ideas concerning biogeographic patterns and their causes. Alfred Russel Wallace was the first to recognize a division between Asiatic and Australian floras and faunas within the region, and in 1863 he drew a boundary between the islands of Bali and Lombok and farther to the north between Borneo and Sulawesi, at the Makassar Strait. The site of this boundary, now known as Wallace's Line, and its significance for different organisms, has been the subject of much discussion since it was first proposed, and indeed Wallace himself proposed different positions for the boundary at different times. Later authors drew boundaries in other places; and it is now recognized that there is a biogeographic region of transition, named Wallacea, situated between areas with Asiatic and Australian floras and faunas, with elements of both but where organisms show a high degree of endemism. This region extends from east of the Makassar Strait to west of the Bird's Head of New Guinea. *See* BIOGEOGRAPHY.

The origin of the biogeographic patterns is complex. There was dispersal of species throughout the region as fragments of Gondwana collided with Asia and as new land and oceanographic connections were established when sea level changed. Vicariance (fragmentation of ancestral species by newly created physical or ecological barriers) resulted from the rise of mountains, formation of deep marine basins, changes in distribution of land and sea as sea levels changed, and modification of habitats as climates changed. The tropics are regions of immense genetic diversity, and there was also rapid speciation as new habitats were established due to emergence of new landmasses. New Guinea is one example of such a pool of diversity. Mountains have risen rapidly in northern New Guinea in the last few million years as the result of collision of Pacific island arcs with the north Australian margin. As a result, climate changed, new habitats were created, and as new barriers separated some species, new pathways developed for the dispersal of others. Many animals and plants of Australian and Pacific origin speciated rapidly in this newly emergent region.

In addition, the tectonic convergence in the region provided new connections between Australia and Asia significantly modified by global climatic changes in the last million years. Glacially related lowering of global sea level resulted in emergence of the Sunda Shelf, and temporary land bridges joined present-day Sumatra, Java, Bali, and Borneo to mainland Asia, allowing the exchange of animals and plants. Major parts of the Sahul Shelf and Arafura Shelf were also emergent, allowing dispersal from Australia into New Guinea and surrounding islands. During interglacial periods the climate warmed again, sea level rose, and islands were once again separated. This process was repeated several times. However, many islands of Wallacea remained isolated by deep water which formed an effective barrier to the migration of most plants and animals between Asia and Australia. Local fluctuations in rainfall and temperature, and consequently vegetation, enhanced the effects of oscillations of dispersal and vicariance and left pockets of isolated fauna and flora. Thus, the plants and animals of Sundaland remain closely related to those of Asia, whereas those of northern Australia, New Guinea, and surrounding islands are predominantly Australian with Pacific influences. Wallace's Line should not be expected to be a sharp single boundary for all organisms at one place. The biogeographic region of Wallacea is a moving interface between Asian and Australian flora and fauna reflecting the establishment of new migration pathways about 20 million years ago. These new pathways were due to collision, the barriers to complete intermingling due to deep marine basins, the effects of glacial climatic changes, differences in abilities of different organisms to disperse, and complications owing to possible remnants of Gondwana elements derived from both Asian and Australian sides.

Resources. The entire region is immensely rich in natural resources, in particular petroleum, minerals, and timber. Petroleum has been produced in large quantities on land in Sumatra, Borneo, and the Bird's Head of New Guinea since the mid-twentieth century. Most oil and gas provinces are in Cenozoic basins, and in the late 1990s the East Indies provided about 5% of annual world oil production. In recent years, exploration and production has moved offshore and is increasingly moving into deeper waters. Many parts of the region have considerable potential for geothermal energy production. Mineral production has been historically important, with major discoveries of tin, gold, and nickel. The late-twentieth-century discovery of major mineral deposits in the northern New Guinea margin suggests that the young island arcs of the region will continue to be targets for exploration, and large deposits are likely to be found both on and off shore. *See* EARTH RESOURCE PATTERNS.

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Eating disorders

Eating disorders are characterized by abnormal eating behaviors and beliefs about eating, weight, and body shape. The two formal diagnoses are anorexia nervosa and bulimia nervosa. In addition, there are many cases of abnormal eating that have only some of the features required for an eating disorder diagnosis; these cases are classified as eating disorders not otherwise specified. Another eating disorder diagnosis has been proposed, termed binge eating disorder. Obesity is classified as a general medical condition and not as an eating disorder (a psychiatric condition) because it is not consistently associated with psychological or behavioral problems.

There are also three childhood eating disorders: (1) Pica is a persistent pattern of eating nonnutritive substances in infants or young children. (2) Rumination disorder involves repeated regurgitation and rechewing of food. This behavior is not the result of a gastrointestinal or medical condition; the partially digested food comes back into the mouth without any observable nausea, disgust, or attempt to vomit. (3) Feeding disorder of infancy or early childhood is the persistent failure to eat adequately, as reflected in insufficient weight gain for age. Pica and rumination disorder are thought to be uncommon and frequently associated with developmental delays and mental retardation. Perhaps half of the pediatric hospitalizations for inadequate weight gain (which constitute 1–5% of all pediatric hospitalizations) may be due to feeding disorder of infancy or early childhood.

Anorexia nervosa. Anorexia nervosa is characterized by a refusal to maintain a minimal normal body weight (defined as 15% below average weight for height), an intense fear of becoming fat, and, if female, amenorrhea for at least 3 months. The majority of cases of anorexia nervosa are classified as restricting type; these individuals achieve abnormally low weight by severely dieting, fasting, and often by exercising compulsively. In severe cases, patients refuse to eat and can die of starvation or severe medical complications. Another subtype of anorexia nervosa is binge eating/purging type; in addition to the low weight, binge eating and purging through vomiting or the misuse of laxatives, enemas, or diuretics occur. Despite being emaciated or dangerously thin, persons with anorexia nervosa perceive themselves as overweight (distorted body image), deny the seriousness of their condition, and have an intense fear of becoming fat. This fear often increases even as weight continues to drop. Self-evaluation is dominated by the importance of weight and body shape.

Epidemiology and associated features. Anorexia nervosa occurs in roughly 1% of adolescent and young adult females. Most cases (90%) are female, and the majority are Caucasian and come from middle-class or higher socioeconomic groups. This disorder is more prevalent in industrialized countries that share western views regarding thinness as an ideal. It develops most frequently during adolescence; mean age of onset is 17 years with bimodal peaks at ages 14 and 18. The prevalence of anorexia nervosa has remained constant over the past few decades; one notable change is the higher incidence in women over age 30, although this still represents a minority of cases.

Persons with anorexia nervosa frequently manifest symptoms of depression (for example, depressed mood, loss of interest, social withdrawal, insomnia, difficulty in concentrating) and anxiety. It is sometimes difficult to ascertain whether these symptoms reflect secondary effects of the starvation or the coexistence of psychiatric disorders. The restricting type of anorexia nervosa is associated with obsessionality, rigidity, perfectionism, and overcontrol, whereas the binge/purge subtype is associated with greater mood instability and impulsivity across a wide range of areas, including substance abuse.

Although some cases of anorexia nervosa show no evidence of medical problems, prolonged starvation affects most organ systems, and a wide array of medical problems tend to be present. Frequent medical findings include leukopenia (decrease in white blood cells) and anemia, dehydration, hypercholesterolemia, abnormal liver and thyroid functions, electrolyte imbalances, low estrogen levels (females), osteoporosis, and cardiovascular problems (sinus bradycardia, severe hypotension, and arrhythmias). Anorexia nervosa tends to be a chronic refractory condition and is associated with considerable morbidity and mortality. Long-term mortality from anorexia nervosa is estimated at 5–10% with most deaths resulting from starvation, cardiac events, or suicide.

Etiology and maintenance. The causes of anorexia nervosa are not yet understood but are likely to involve a complex combination of genetic, familial, psychological, and sociocultural factors. There is an increased risk of anorexia nervosa among first-degree relatives of persons with the disorder and perhaps with depression. Individuals with certain psychological styles (low self-esteem, perfectionism, obsessiveness) and from families with certain characteristics (rigidity, overinvolvement, poor communication, high achievement demands) may be particularly susceptible to western society's pervasive emphasis on thinness as the ideal. The onset of anorexia nervosa tends to follow a period of dieting and is frequently triggered by a stressful life event or transition.

Treatment. Since the starvation and weight loss can be life-threatening, initial treatment efforts need to focus on weight gain and the reestablishment of regular eating patterns. Inpatient hospitalization is frequently necessary to prevent continued weight loss, to stabilize the weight and any medical comorbidities, and to initiate the nutritional rehabilitation. The initial period of weight stabilization and nutritional rehabilitation is typically met by the patient with great resistance and emotional upset. Consistency and structure during this period are essential

for the patient's success, and family therapy is generally helpful to family members.

Although significant psychological issues tend to be present, it is generally ineffective to address these until weight has been stabilized. The effects of starvation on psychological functioning coupled with the obsessional preoccupation with food and weight preclude any meaningful psychotherapy. Once weight gain is achieved, psychotherapies can become useful. Available evidence favors the use of cognitive-behavioral therapy and family-based therapy. Effective medication treatments have not been found. Relapse rates are high, and relapse is most common during the year following treatment, although it can occur even years later. Patients who achieve normal weight during treatment and are able to maintain that weight have the best long-term prognosis. *See* ANOREXIA NERVOSA; PSYCHOTHERAPY.

Bulimia nervosa. Bulimia nervosa is characterized by recurrent episodes of binge eating (eating large amounts of food while experiencing a subjective sense of lack of control over the eating), the regular use of extreme weight compensatory methods (for example, self-induced vomiting; abuse of laxatives, diuretics, or diet pills; severe dieting or fasting; vigorous exercise), and dysfunctional beliefs about weight and shape that unduly influence self-evaluation or self-worth.

Epidemiology and associated features. Bulimia nervosa occurs in roughly 2% of adolescents and adults. It is most common in females (90% of cases), Caucasians, and middle-class or higher socioeconomic groups. The prevalence of bulimia has increased over the past few decades; persons born after 1960 have a higher rate of developing bulimia. Bulimia nervosa is also becoming more common in non-Caucasian groups.

Persons with bulimia nervosa have high rates of depression, anxiety, and substance abuse problems. The majority of persons with bulimia nervosa are of average weight, although frequent weight fluctuations are common. The preoccupation with weight and shape coupled with considerable shame and guilt regarding their eating practices contributes to frequent interpersonal and social difficulties.

Although this condition is less dangerous than anorexia nervosa, medical complications can occur in bulimia nervosa because of the extreme dieting, binge eating, and purging. Dental erosion and periodontal problems are common. Electrolyte imbalance and dehydration can result in serious medical complications, including cardiac arrhythmias. In rare cases, esophageal bleeding and gastric ruptures occur.

Etiology and maintenance. Bulimia nervosa is likely to result from a combination of genetic, familial, psychological, and sociocultural factors. A family history of bulimia and obesity is considered a risk factor for developing bulimia. In addition, negative experiences such as teasing about weight and shape may precipitate dieting; strict dieting is thought to represent a major risk factor for the development of bulimia. Although many persons have weight and diet concerns, the development of bulimia is thought to arise only in vulnerable individuals and usually after a stressful event. Bulimia nervosa is a self-maintaining vicious cycle: strict dieting and deprivation triggers binge eating, followed by purging to undo the caloric intake and emotional fear of weight gain, followed by shame and disgust, followed by strict dieting.

Treatment. Bulimia nervosa can often be treated successfully with outpatient therapies by appropriately trained professionals. Cognitive-behavioral therapy and interpersonal psychotherapy have been found to be most effective for reducing binge eating and vomiting and improving associated concerns such as depression, self-esteem, and body image. These two therapies also have the best results over the long term. Certain types of pharmacotherapy, notably antidepressant medications, are also effective, although their benefits are less certain once the medication is discontinued. Antidepressant medications appear to help people with bulimia nervosa even if they are not depressed. The combination of antidepressants and cognitive-behavioral therapy appears to be superior to medication treatment alone but not to cognitive-behavioral therapy alone.

Binge eating disorder. Binge eating disorder is characterized by recurrent episodes of binge eating but, unlike bulimia nervosa, no extreme weight control behaviors (purging, laxatives, fasting) are present. Persons with binge eating disorder have chaotic eating patterns and frequently overeat as well as binge.

Epidemiology and associated features. Although obesity is not required for the diagnosis, many people with binge eating disorder are overweight, because strict dieting and frequent purging are not present. Binge eating disorder is estimated to occur in 3% of the general population but in roughly 30% of obese persons. Binge eating disorder occurs most frequently in adulthood, affects men nearly as often as women, and occurs across different ethnic groups.

Overweight persons with binge eating disorder differ from overweight persons who do not binge. In particular, obese binge eaters are characterized by higher levels of psychiatric problems (depression, anxiety, substance abuse) and psychological problems (poor self-esteem, body image dissatisfaction) than non–binge eaters and closely resemble persons with bulimia nervosa. Overweight persons with binge eating disorder are at high risk for further weight gain and weight fluctuations and associated medical complications.

Etiology. The etiology of binge eating disorder is unknown. Adverse childhood experiences, family histories of depression and obesity, and teasing experiences about weight and shape have been identified as risk factors. In contrast to the other eating disorders, a history of dieting does not seem to predate binge eating in roughly half of the cases.

Treatment. Cognitive-behavioral therapy is effective for reducing binge eating and improving associated concerns such as depression, self-esteem, and body image, but does not seem to result in weight loss. There is some evidence that behavioral weight control treatment can reduce binge eating and facilitate

weight loss. Antidepressant medications appear to reduce binge eating but do not produce weight loss; relapse is rapid after discontinuation of the medication. *See* AFFECTIVE DISORDERS. Carlos M. Grilo

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