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Eye (anatomy)

I –INTRODUCTION

Eye (anatomy), light-sensitive organ of vision in animals. The eyes of various species vary from simple structures that are capable only of differentiating between light and dark to complex organs, such as those of humans and other mammals, that can distinguish minute variations of shape, color, brightness, and distance. The actual process of seeing is performed by the brain rather than by the eye. The function of the eye is to translate the electromagnetic vibrations of light into patterns of nerve impulses that are transmitted to the brain.

II -THE HUMAN EYE

The entire eye, often called the eyeball, is a spherical structure approximately 2.5 cm (about 1 in) in diameter with a pronounced bulge on its forward surface. The outer part of the eye is composed of three layers of tissue. The outside layer is the sclera, a protective coating. It covers about five-sixths of the surface of the eye. At the front of the eyeball, it is continuous with the bulging, transparent cornea. The middle layer of the coating of the eye is the choroid, a vascular layer lining the posterior three-fifths of the eyeball. The choroid is continuous with the ciliary body and with the iris, which lies at the front of the eye. The innermost layer is the light-sensitive retina.

The cornea is a tough, five-layered membrane through which light is admitted to the interior of the eye. Behind the cornea is a chamber filled with clear, watery fluid, the aqueous humor, which separates the cornea from the crystalline lens. The lens itself is a flattened sphere constructed of a large number of transparent fibers arranged in layers. It is connected by ligaments to a ringlike muscle, called the ciliary muscle, which surrounds it. The ciliary muscle and its surrounding tissues form the ciliary body. This muscle, by flattening the lens or making it more nearly spherical, changes its focal length.

The pigmented iris hangs behind the cornea in front of the lens, and has a circular opening in its center. The size of its opening, the pupil, is controlled by a muscle around its edge. This muscle contracts or relaxes, making the pupil larger or smaller, to control the amount of light admitted to the eye.

Behind the lens the main body of the eye is filled with a transparent, jellylike substance, the vitreous humor, enclosed in a thin sac, the hyaloid membrane. The pressure of the vitreous humor keeps the eyeball distended.

The retina is a complex layer, composed largely of nerve cells. The light-sensitive receptor cells lie on the outer surface of the retina in front of a pigmented tissue layer. These cells take the form of rods or cones packed closely together like matches in a box. Directly behind the pupil is a small yellow-pigmented spot, the macula lutea, in the center of which is the fovea centralis, the area of greatest visual acuity of the eye. At the center of the fovea, the sensory layer is composed entirely of cone-shaped cells. Around the fovea both rod-shaped and cone-shaped cells are present, with the cone-shaped cells becoming fewer toward the periphery of the sensitive area. At the outer edges are only rod-shaped cells.

Where the optic nerve enters the eyeball, below and slightly to the inner side of the fovea, a small round area of the retina exists that has no light-sensitive cells. This optic disk forms the blind spot of the eye.

III -FUNCTIONING OF THE EYE

In general the eyes of all animals resemble simple cameras in that the lens of the eye forms an inverted image of objects in front of it on the sensitive retina, which corresponds to the film in a camera.

Focusing the eye, as mentioned above, is accomplished by a flattening or thickening (rounding) of the lens. The process is known as accommodation. In the normal eye accommodation is not necessary for seeing distant objects. The lens, when flattened by the suspensory ligament, brings such objects to focus on the retina. For nearer objects the lens is increasingly rounded by ciliary muscle contraction, which relaxes the suspensory ligament. A young child can see clearly at a distance as close as 6.3 cm (2.5 in), but with increasing age the lens gradually hardens, so that the limits of close seeing are approximately 15 cm (about 6 in) at the age of 30 and 40 cm (16 in) at the age of 50. In the later years of life most people lose the ability to accommodate their eyes to distances within reading or close working range. This condition, known as presbyopia, can be corrected by the use of special convex lenses for the near range.

Structural differences in the size of the eye cause the defects of hyperopia, or farsightedness, and myopia, or nearsightedness. See Eyeglasses; Vision.

As mentioned above, the eye sees with greatest clarity only in the region of the fovea; due to the neural structure of the retina. The cone-shaped cells of the retina are individually connected to other nerve fibers, so that stimuli to each individual cell are reproduced and, as a result, fine details can be distinguished. The rodshaped cells, on the other hand, are connected in groups so that they respond to stimuli over a general area.

The rods, therefore, respond to small total light stimuli, but do not have the ability to separate small details of the visual image. The result of these differences in structure is that the visual field of the eye is composed of a small central area of great sharpness surrounded by an area of lesser sharpness. In the latter area, however, the sensitivity of the eye to light is great. As a result, dim objects can be seen at night on the peripheral part of the retina when they are invisible to the central part.

The mechanism of seeing at night involves the sensitization of the rod cells by means of a pigment, called visual purple or rhodopsin, that is formed within the cells. Vitamin A is necessary for the production of visual purple; a deficiency of this vitamin leads to night blindness. Visual purple is bleached by the action of light and must be reformed by the rod cells under conditions of darkness. Hence a person who steps from sunlight into a darkened room cannot see until the pigment begins to form. When the pigment has formed and the eyes are sensitive to low levels of illumination, the eyes are said to be dark-adapted.

A brownish pigment present in the outer layer of the retina serves to protect the cone cells of the retina from overexposure to light. If bright light strikes the retina, granules of this brown pigment migrate to the spaces around the cone cells, sheathing and screening them from the light. This action, called light adaptation, has the opposite effect to that of dark adaptation.

Subjectively, a person is not conscious that the visual field consists of a central zone of sharpness surrounded by an area of increasing fuzziness.

The reason is that the eyes are constantly moving, bringing first one part of the visual field and then another to the foveal region as the attention is shifted from one object to another. These motions are accomplished by six muscles that move the eyeball upward, downward, to the left, to the right, and obliquely. The motions of the eye muscles are extremely precise; the estimation has been made that the eyes can be moved to focus on no less than 100,000 distinct points in the visual field. The muscles of the two eyes, working together, also serve the important function of converging the eyes on any point being observed, so that the images of the two eyes coincide. When convergence is nonexistent or faulty, double vision results. The movement of the eyes and fusion of the images also play a part in the visual estimation of size and distance.

IV -PROTECTIVE STRUCTURES

Several structures, not parts of the eyeball, contribute to the protection of the eye. The most important of these are the eyelids, two folds of skin and tissue, upper and lower, that can be closed by means of muscles to form a protective covering over the eyeball against excessive light and mechanical injury.

The eyelashes, a fringe of short hairs growing on the edge of either eyelid, act as a screen to keep dust particles and insects out of the eyes when the eyelids are partly closed. Inside the eyelids is a thin protective membrane, the conjunctiva, which doubles over to cover the visible sclera. Each eye also has a tear gland, or lacrimal organ, situated at the outside corner of the eye. The salty secretion of these glands lubricates the forward part of the eyeball when the eyelids are closed and flushes away any small dust particles or other foreign matter on the surface of the eye. Normally the eyelids of human eyes close by reflex action about every six seconds, but if dust reaches the surface of the eye and is not washed away, the eyelids blink oftener and more tears are produced. On the edges of the eyelids are a number of small glands, the Meibomian glands, which produce a fatty secretion that lubricates the eyelids themselves and the eyelashes. The eyebrows, located above each eye, also have a protective function in soaking up or deflecting perspiration or rain and preventing the moisture from running into the eyes. The hollow socket in the skull in which the eye is set is called the orbit. The bony edges of the orbit, the frontal bone, and the cheekbone protect the eye from mechanical injury by blows or collisions.

V -COMPARATIVE ANATOMY

The simplest animal eyes occur in the cnidarians and ctenophores, phyla comprising the jellyfish and somewhat similar primitive animals. These eyes, known as pigment eyes, consist of groups of pigment cells associated with sensory cells and often covered with a thickened layer of cuticle that forms a kind of lens. Similar eyes, usually having a somewhat more complex structure, occur in worms, insects, and mollusks.

Two kinds of image-forming eyes are found in the animal world, single and compound eyes. The single eyes are essentially similar to the human eye, though varying from group to group in details of structure. The lowest species to develop such eyes are some of the large jellyfish. Compound eyes, confined to the arthropods (see Arthropod), consist of a faceted lens, each facet of which forms a separate image on a retinal cell, creating a moasic field. In some arthropods the structure is more sophisticated, forming a combined image.

The eyes of other vertebrates are essentially similar to human eyes, although important modifications may exist. The eyes of such nocturnal animals as cats, owls, and bats are provided only with rod cells, and the cells are both more sensitive and more numerous than in humans. The eye of a dolphin has 7000 times as many rod cells as a human eye, enabling it to see in deep water. The eyes of most fish have a flat cornea and a globular lens and are hence particularly adapted for seeing close objects. Birds' eyes are elongated from front to back, permitting larger images of distant objects to be formed on the retina.

VI -EYE DISEASES

Eye disorders may be classified according to the part of the eye in which the disorders occur.

The most common disease of the eyelids is hordeolum, known commonly as a sty, which is an infection of the follicles of the eyelashes, usually caused by infection by staphylococci. Internal sties that occur inside the eyelid and not on its edge are similar infections of the lubricating Meibomian glands. Abscesses of the eyelids are sometimes the result of penetrating wounds. Several congenital defects of the eyelids occasionally occur, including coloboma, or cleft eyelid, and ptosis, a drooping of the upper lid. Among acquired defects are symblepharon, an adhesion of the inner surface of the eyelid to the eyeball, which is most frequently the result of burns. Entropion, the turning of the eyelid inward toward the cornea, and ectropion, the turning of the eyelid outward, can be caused by scars or by spasmodic muscular contractions resulting from chronic irritation.

The eyelids also are subject to several diseases of the skin such as eczema and acne, and to both benign and malignant tumors. Another eye disease is infection of the conjunctiva, the mucous membranes covering the inside of the eyelids and the outside of the eyeball. See Conjunctivitis; Trachoma.

Disorders of the cornea, which may result in a loss of transparency and impaired sight, are usually the result of injury but may also occur as a secondary result of disease; for example, edema, or swelling, of the cornea sometimes accompanies glaucoma.

The choroid, or middle coat of the eyeball, contains most of the blood vessels of the eye; it is often the site of secondary infections from toxic conditions and bacterial infections such as tuberculosis and syphilis. Cancer may develop in the choroidal tissues or may be carried to the eye from malignancies elsewhere in the body. The light-sensitive retina, which lies just beneath the choroid, also is subject to the same type of infections. The cause of retrolental fibroplasia, however—a disease of premature infants that causes retinal detachment and partial blindness—is unknown. Retinal detachment may also follow cataract surgery. Laser beams are sometimes used to weld detached retinas back onto the eye. Another retinal condition, called macular degeneration, affects the central retina. Macular degeneration is a frequent cause of loss of vision in older persons. Juvenile forms of this condition also exist.

The optic nerve contains the retinal nerve fibers, which carry visual impulses to the brain. The retinal circulation is carried by the central artery and vein, which lie in the optic nerve. The sheath of the optic nerve communicates with the cerebral lymph spaces. Inflammation of that part of the optic nerve situated within the eye is known as optic neuritis, or papillitis; when inflammation occurs in the part of the optic nerve behind the eye, the disease is called retrobulbar neuritis. When the pressure in the skull is elevated, or increased in intracranial pressure, as in brain tumors, edema and swelling of the optic disk occur where the nerve enters the eyeball, a condition known as papilledema, or chocked disk.

For disorders of the crystalline lens, see Cataract. See also Color Blindness.

VII -EYE BANK

Eye banks are organizations that distribute corneal tissue taken from deceased persons for eye grafts. Blindness caused by cloudiness or scarring of the cornea can sometimes be cured by surgical removal of the affected portion of the corneal tissue. With present techniques, such tissue can be kept alive for only 48 hours, but current experiments in preserving human corneas by freezing give hope of extending its useful life for months. Eye banks also preserve and distribute vitreous humor, the liquid within the larger chamber of the eye, for use in treatment of detached retinas. The first eye bank was opened in New York City in 1945. The Eye-Bank Association of America, in Rochester, New York, acts as a clearinghouse for information.

Fingerprinting

I –INTRODUCTION

Fingerprinting, method of identification using the impression made by the minute ridge formations or patterns found on the fingertips. No two persons have exactly the same arrangement of ridge patterns, and the patterns of any one individual remain unchanged through life. To obtain a set of fingerprints, the ends of the fingers are inked and then pressed or rolled one by one on some receiving surface. Fingerprints may be classified and filed on the basis of the ridge patterns, setting up an identification system that is almost infallible.

II –HISTORY

The first recorded use of fingerprints was by the ancient Assyrians and Chinese for the signing of legal documents. Probably the first modern study of fingerprints was made by the Czech physiologist Johannes Evengelista Purkinje, who in 1823 proposed a system of classification that attracted little attention. The use of fingerprints for identification purposes was proposed late in the 19th century by the British scientist Sir Francis Galton, who wrote a detailed study of fingerprints in which he presented a new classification system using prints of all ten fingers, which is the basis of identification systems still in use. In the 1890s the police in Bengal, India, under the British police official Sir Edward Richard Henry, began using fingerprints to identify criminals. As assistant commissioner of metropolitan police, Henry established the first British fingerprint files in London in 1901. Subsequently, the use of fingerprinting as a means for identifying criminals spread rapidly throughout Europe and the United States, superseding the old Bertillon system of identification by means of body measurements.

III -MODERN USE

As crime-detection methods improved, law enforcement officers found that any smooth, hard surface touched by a human hand would yield fingerprints made by the oily secretion present on the skin. When these so-called latent prints were dusted with powder or chemically treated, the identifying fingerprint pattern could be seen and photographed or otherwise preserved. Today, law enforcement agencies can also use computers to digitally record fingerprints and to transmit them electronically to other agencies for comparison. By comparing fingerprints at the scene of a crime with the fingerprint record of suspected persons, officials can establish absolute proof of the presence or identity of a person.

The confusion and inefficiency caused by the establishment of many separate fingerprint archives in the United States led the federal government to set up a central agency in 1924, the Identification Division of the Federal Bureau of Investigation (FBI). This division was absorbed in 1993 by the FBI's Criminal Justice Information Services Division, which now maintains the world's largest fingerprint collection. Currently the FBI has a library of more than 234 million civil and criminal fingerprint cards, representing 81 million people. In 1999 the FBI began full operation of the Integrated Automated Fingerprint Identification System (IAFIS), a computerized system that stores digital images of fingerprints for more than 36 million individuals, along with each individual's criminal history if one exists. Using IAFIS, authorities can conduct automated searches to identify people from their fingerprints and determine whether they have a criminal record. The system also gives state and local law enforcement agencies the ability to electronically transmit fingerprint information to the FBI. The implementation of IAFIS represented a breakthrough in crimefighting by reducing the time needed for fingerprint identification from weeks to minutes or hours.

Infrared Radiation

Infrared Radiation, emission of energy as electromagnetic waves in the portion of the spectrum just beyond the limit of the red portion of visible radiation (see Electromagnetic Radiation). The wavelengths of infrared radiation are shorter than those of radio waves and longer than those of light waves. They range between approximately 10-6 and 10-3 (about 0.0004 and 0.04 in).

Infrared radiation may be detected as heat, and instruments such as bolometers are used to detect it. See Radiation; Spectrum.

Infrared radiation is used to obtain pictures of distant objects obscured by atmospheric haze, because visible light is scattered by haze but infrared radiation is not. The detection of infrared radiation is used by astronomers to observe stars and nebulas that are invisible in ordinary light or that emit radiation in the infrared portion of the spectrum.

An opaque filter that admits only infrared radiation is used for very precise infrared photographs, but an ordinary orange or light-red filter, which will absorb blue and violet light, is usually sufficient for most infrared pictures. Developed about 1880, infrared photography has today become an important diagnostic tool in medical science as well as in agriculture and industry. Use of infrared techniques reveals pathogenic conditions that are not visible to the eye or recorded on X-ray plates. Remote sensing by means of aerial and orbital infrared photography has been used to monitor crop conditions and insect and disease damage to large agricultural areas, and to locate mineral deposits. See Aerial Survey; Satellite, Artificial. In industry, infrared spectroscopy forms an increasingly important part of metal and alloy research, and infrared photography is used to monitor the quality of products. See also Photography: Photographic Films.

Infrared devices such as those used during World War II enable sharpshooters to see their targets in total visual darkness. These instruments consist essentially of an infrared lamp that sends out a beam of infrared radiation, often referred to as black light, and a telescope receiver that picks up returned radiation from the object and converts it to a visible image.

Deoxyribonucleic Acid

I –INTRODUCTION

Deoxyribonucleic Acid (DNA), genetic material of all cellular organisms and most viruses. DNA carries the information needed to direct protein synthesis and replication. Protein synthesis is the production of the proteins needed by the cell or virus for its activities and development. Replication is the process by which DNA copies itself for each descendant cell or virus, passing on the information needed for protein synthesis. In most cellular organisms, DNA is organized on chromosomes located in the nucleus of the cell

II –STRUCTURE

A molecule of DNA consists of two chains, strands composed of a large number of chemical compounds, called nucleotides, linked together to form a chain. These chains are arranged like a ladder that has been twisted into the shape of a winding staircase, called a double helix. Each nucleotide consists of three units: a sugar molecule called deoxyribose, a phosphate group, and one of four different nitrogen-containing compounds called bases. The four bases are adenine (A), guanine (G), thymine (T), and cytosine (C). The deoxyribose molecule occupies the center position in the nucleotide, flanked by a phosphate group on one side and a base on the other. The phosphate group of each nucleotide is also linked to the deoxyribose of the adjacent nucleotide in the chain. These linked deoxyribose-phosphate subunits form the parallel side rails of the ladder. The bases face inward toward each other, forming the rungs of the ladder.

The nucleotides in one DNA strand have a specific association with the corresponding nucleotides in the other DNA strand. Because of the chemical affinity of the bases, nucleotides containing adenine are always paired with nucleotides containing thymine, and nucleotides containing cytosine are always paired with nucleotides containing guanine. The complementary bases are joined to each other by weak chemical bonds called hydrogen bonds.

In 1953 American biochemist James D. Watson and British biophysicist Francis Crick published the first description of the structure of DNA. Their model proved to be so important for the understanding of protein synthesis, DNA replication, and mutation that they were awarded the 1962 Nobel Prize for physiology or medicine for their work.

III -PROTEIN SYNTHESIS

DNA carries the instructions for the production of proteins. A protein is composed of smaller molecules called amino acids, and the structure and function of the protein is determined by the sequence of its amino acids. The sequence of amino acids, in turn, is determined by the sequence of nucleotide bases in the DNA. A sequence of three nucleotide bases, called a triplet, is the genetic code word, or codon, that specifies a particular amino acid. For instance, the triplet GAC (guanine, adenine, and cytosine) is the codon for the amino acid leucine, and the triplet CAG (cytosine, adenine, and guanine) is the codon for the amino acid valine. A protein consisting of 100 amino acids is thus encoded by a DNA segment consisting of 300 nucleotides. Of the two polynucleotide chains that form a DNA molecule, only one strand contains the information needed for the production of a given amino acid sequence. The other strand aids in replication.

Protein synthesis begins with the separation of a DNA molecule into two strands. In a process called transcription, a section of one strand acts as a template, or pattern, to produce a new strand called messenger RNA (mRNA). The mRNA leaves the cell nucleus and attaches to the ribosomes, specialized cellular structures that are the sites of protein synthesis. Amino acids are carried to the ribosomes by another type of RNA, called transfer RNA (tRNA). In a process called translation, the amino acids are linked together in a particular sequence, dictated by the mRNA, to form a protein.

A gene is a sequence of DNA nucleotides that specify the order of amino acids in a protein via an intermediary mRNA molecule. Substituting one DNA nucleotide with another containing a different base causes all descendant cells or viruses to have the altered nucleotide base sequence. As a result of the substitution, the sequence of amino acids in the resulting protein may also be changed. Such a change in a DNA molecule is called a mutation. Most mutations are the result of errors in the replication process. Exposure of a cell or virus to radiation or to certain chemicals increases the likelihood of mutations.

IV –REPLICATION

In most cellular organisms, replication of a DNA molecule takes place in the cell nucleus and occurs just before the cell divides. Replication begins with the separation of the two polynucleotide chains, each of which then acts as a template for the assembly of a new complementary chain. As the old chains separate, each nucleotide in the two chains attracts a complementary nucleotide that has been formed earlier by the cell. The nucleotides are joined to one another by hydrogen bonds to form the rungs of a new DNA molecule. As the complementary nucleotides are fitted into place, an enzyme called DNA polymerase links them together by bonding the phosphate group of one nucleotide to the sugar molecule of the adjacent nucleotide, forming the side rail of the new DNA molecule. This process continues until a new polynucleotide chain has been formed alongside the old one, forming a new double-helix molecule.

V -TOOLS AND PROCEDURES

Several tools and procedures facilitate are used by scientists for the study and manipulation of DNA. Specialized enzymes, called restriction enzymes, found in bacteria act like molecular scissors to cut the phosphate backbones of DNA molecules at specific base sequences. Strands of DNA that have been cut with restriction enzymes are left with single-stranded tails that are called sticky ends, because they can easily realign with tails from certain other DNA fragments. Scientists take advantage of restriction enzymes and the sticky ends generated by these enzymes to carry out recombinant DNA technology, or genetic engineering. This technology involves removing a specific gene from one organism and inserting the gene into another organism.

Another tool for working with DNA is a procedure called polymerase chain reaction (PCR). This procedure uses the enzyme DNA polymerase to make copies of DNA strands in a process that mimics the way in which DNA replicates naturally within cells. Scientists use PCR to obtain vast numbers of copies of a given segment of DNA.

DNA fingerprinting, also called DNA typing, makes it possible to compare samples of DNA from various sources in a manner that is analogous to the comparison of fingerprints. In this procedure, scientists use restriction enzymes to cleave a sample of DNA into an assortment of fragments. Solutions containing these fragments are placed at the surface of a gel to which an electric current is applied. The electric current causes the DNA fragments to move through the gel. Because smaller fragments move more quickly than larger ones, this process, called electrophoresis, separates the fragments according to their size. The fragments are then marked with probes and exposed on X-ray film, where they form the DNA fingerprint—a pattern of characteristic black bars that is unique for each type of DNA.

A procedure called DNA sequencing makes it possible to determine the precise order, or sequence, of nucleotide bases within a fragment of DNA. Most versions of DNA sequencing use a technique called primer extension, developed by British molecular biologist Frederick Sanger.

In primer extension, specific pieces of DNA are replicated and modified, so that each DNA segment ends in a fluorescent form of one of the four nucleotide bases. Modern DNA sequencers, pioneered by American molecular biologist Leroy Hood, incorporate both lasers and computers. Scientists have completely sequenced the genetic material of several microorganisms, including the bacterium Escherichia coli. In 1998, scientists achieved the milestone of sequencing the complete genome of a multicellular organism—a roundworm identified as Caenorhabditis elegans. The Human Genome Project, an international research collaboration, has been established to determine the sequence of all of the three billion nucleotide base pairs that make up the human genetic material.

An instrument called an atomic force microscope enables scientists to manipulate the three-dimensional structure of DNA molecules. This microscope involves laser beams that act like tweezers—attaching to the ends of a DNA molecule and pulling on them. By manipulating these laser beams, scientists can stretch, or uncoil, fragments of DNA. This work is helping reveal how DNA changes its three-dimensional shape as it interacts with enzymes.

VI –APPLICATIONS

Research into DNA has had a significant impact on medicine. Through recombinant DNA technology, scientists can modify microorganisms so that they become so-called factories that produce large quantities of medically useful drugs. This technology is used to produce insulin, which is a drug used by diabetics, and interferon, which is used by some cancer patients. Studies of human DNA are revealing genes that are associated with specific diseases, such as cystic fibrosis and breast cancer. This information is helping physicians to diagnose various diseases, and it may lead to new treatments. For example, physicians are using a technology called chimeraplasty, which involves a synthetic molecule containing both DNA and RNA strands, in an effort to develop a treatment for a form of hemophilia.

Forensic science uses techniques developed in DNA research to identify individuals who have committed crimes. DNA from semen, skin, or blood taken from the crime scene can be compared with the DNA of a suspect, and the results can be used in court as evidence.

DNA has helped taxonomists determine evolutionary relationships among animals, plants, and other life forms. Closely related species have more similar DNA than do species that are distantly related. One surprising finding to emerge from DNA studies is that vultures of the Americas are more closely related to storks than to the vultures of Europe, Asia, or Africa (see Classification).

Techniques of DNA manipulation are used in farming, in the form of genetic engineering and biotechnology. Strains of crop plants to which genes have been transferred may produce higher yields and may be more resistant to insects. Cattle have been similarly treated to increase milk and beef production, as have hogs, to yield more meat with less fat.

VII -SOCIAL ISSUES

Despite the many benefits offered by DNA technology, some critics argue that its development should be monitored closely. One fear raised by such critics is that DNA fingerprinting could provide a means for employers to discriminate against members of various ethnic groups. Critics also fear that studies of people's DNA could permit insurance companies to deny health insurance to those people at risk for developing certain diseases. The potential use of DNA technology to alter the genes of embryos is a particularly controversial issue.

The use of DNA technology in agriculture has also sparked controversy. Some people question the safety, desirability, and ecological impact of genetically altered crop plants. In addition, animal rights groups have protested against the genetic engineering of farm animals.

Despite these and other areas of disagreement, many people agree that DNA technology offers a mixture of benefits and potential hazards. Many experts also agree that an informed public can help assure that DNA technology is used wisely.

Blood

I –INTRODUCTION

Blood, vital fluid found in humans and other animals that provides important nourishment to all body organs and tissues and carries away waste materials. Sometimes referred to as "the river of life," blood is pumped from the heart through a network of blood vessels collectively known as the circulatory system.

An adult human has about 5 to 6 liters (1 to 2 gal) of blood, which is roughly 7 to 8 percent of total body weight. Infants and children have comparably lower volumes of blood, roughly proportionate to their smaller size. The volume of blood in an individual fluctuates. During dehydration, for example while running a marathon, blood volume decreases. Blood volume increases in circumstances such as pregnancy, when the mother's blood needs to carry extra oxygen and nutrients to the baby.

II -ROLE OF BLOOD

Blood carries oxygen from the lungs to all the other tissues in the body and, in turn, carries waste products, predominantly carbon dioxide, back to the lungs where they are released into the air. When oxygen transport fails, a person dies within a few minutes. Food that has been processed by the digestive system into smaller components such as proteins, fats, and carbohydrates is also delivered to the tissues by the blood. These nutrients provide the materials and energy needed by individual cells for metabolism, or the performance of cellular function. Waste products produced during metabolism, such as urea and uric acid, are carried by the blood to the kidneys, where they are transferred from the blood into urine and eliminated from the body. In addition to oxygen and nutrients, blood also transports special chemicals, called hormones, that regulate certain body functions. The movement of these chemicals enables one organ to control the function of another even though the two organs may be located far apart. In this way, the blood acts not just as a means of transportation but also as a communications system.

The blood is more than a pipeline for nutrients and information; it is also responsible for the activities of the immune system, helping fend off infection and fight disease. In addition, blood carries the means for stopping itself from leaking out of the body after an injury. The blood does this by carrying special cells and proteins, known as the coagulation system, that start to form clots within a matter of seconds after injury.

Blood is vital to maintaining a stable body temperature; in humans, body temperature normally fluctuates within a degree of 37.0° C (98.6° F). Heat production and heat loss in various parts of the body are balanced out by heat transfer via the bloodstream. This is accomplished by varying the diameter of blood vessels in the skin. When a person becomes overheated, the vessels dilate and an increased volume of blood flows through the skin. Heat dissipates through the skin, effectively lowering the body temperature. The increased flow of blood in the skin makes the skin appear pink or flushed. When a person is cold, the skin may become pale as the vessels narrow, diverting blood from the skin and reducing heat loss.

III -COMPOSITION OF BLOOD

About 55 percent of the blood is composed of a liquid known as plasma. The rest of the blood is made of three major types of cells: red blood cells (also known as erythrocytes), white blood cells (leukocytes), and platelets (thrombocytes).

A Plasma

Plasma consists predominantly of water and salts. The kidneys carefully maintain the salt concentration in plasma because small changes in its concentration will cause cells in the body to function improperly. In extreme conditions this can result in seizures, coma, or even death. The pH of plasma, the common measurement of the plasma's acidity, is also carefully controlled by the kidneys within the neutral range of 6.8 to 7.7. Plasma also contains other small molecules, including vitamins, minerals, nutrients, and waste products. The concentrations of all of these molecules must be carefully regulated.

Plasma is usually yellow in color due to proteins dissolved in it. However, after a person eats a fatty meal, that person's plasma temporarily develops a milky color as the blood carries the ingested fats from the intestines to other organs of the body.

Plasma carries a large number of important proteins, including albumin, gamma globulin, and clotting factors. Albumin is the main protein in blood. It helps regulate the water content of tissues and blood. Gamma globulin is composed of tens of thousands of unique antibody molecules. Antibodies neutralize or help destroy infectious organisms. Each antibody is designed to target one specific invading organism. For example, chicken pox antibody will target chicken pox virus, but will leave an influenza virus unharmed. Clotting factors, such as fibrinogen, are involved in forming blood clots that seal leaks after an injury. Plasma that has had the clotting factors removed is called serum. Both serum and plasma are easy to store and have many medical uses.

B -Red Blood Cells

Red blood cells make up almost 45 percent of the blood volume. Their primary function is to carry oxygen from the lungs to every cell in the body. Red blood cells are composed predominantly of a protein and iron compound, called hemoglobin, that captures oxygen molecules as the blood moves through the lungs, giving blood its red color. As blood passes through body tissues, hemoglobin then releases the oxygen to cells throughout the body. Red blood cells are so packed with hemoglobin that they lack many components, including a nucleus, found in other cells.

The membrane, or outer layer, of the red blood cell is flexible, like a soap bubble, and is able to bend in many directions without breaking. This is important because the red blood cells must be able to pass through the tiniest blood vessels, the capillaries, to deliver oxygen wherever it is needed. The capillaries are so narrow that the red blood cells, normally shaped like a disk with a concave top and bottom, must bend and twist to maneuver single file through them.

C -Blood Type

There are several types of red blood cells and each person has red blood cells of just one type. Blood type is determined by the occurrence or absence of substances, known as recognition markers or antigens, on the surface of the red blood cell. Type A blood has just marker A on its red blood cells while type B has only marker B. If neither A nor B markers are present, the blood is type O. If both the A and B markers are present, the blood is type AB. Another marker, the Rh antigen (also known as the Rh factor), is present or absent regardless of the presence of A and B markers. If the Rh marker is present, the blood is said to be Rh positive, and if it is absent, the blood is Rh negative. The most common blood type is A positive—that is, blood that has an A marker and also an Rh marker. More than 20 additional red blood cell types have been discovered.

Blood typing is important for many medical reasons. If a person loses a lot of blood, that person may need a blood transfusion to replace some of the lost red blood cells. Since everyone makes antibodies against substances that are foreign, or not of their own body, transfused blood must be matched so as not to contain these substances. For example, a person who is blood type A positive will not make antibodies against the A or Rh markers, but will make antibodies against the B marker, which is not on that person's own red blood cells. If blood containing the B marker (from types B positive, B negative, AB positive, or AB negative) is transfused into this person, then the transfused red blood cells will be rapidly destroyed by the patient's anti-B antibodies.

In this case, the transfusion will do the patient no good and may even result in serious harm. For a successful blood transfusion into an A positive blood type individual, blood that is type O negative, O positive, A negative, or A positive is needed because these blood types will not be attacked by the patient's anti-B antibodies.

D -White Blood Cells

White blood cells only make up about 1 percent of blood, but their small number belies their immense importance. They play a vital role in the body's immune system—the primary defense mechanism against invading bacteria, viruses, fungi, and parasites. They often accomplish this goal through direct attack, which usually involves identifying the invading organism as foreign, attaching to it, and then destroying it. This process is referred to as phagocytosis.

White blood cells also produce antibodies, which are released into the circulating blood to target and attach to foreign organisms. After attachment, the antibody may neutralize the organism, or it may elicit help from other immune system cells to destroy the foreign substance. There are several varieties of white blood cells, including neutrophils, monocytes, and lymphocytes, all of which interact with one another and with plasma proteins and other cell types to form the complex and highly effective immune system.

E -Platelets and Clotting

The smallest cells in the blood are the platelets, which are designed for a single purpose—to begin the process of coagulation, or forming a clot, whenever a blood vessel is broken. As soon as an artery or vein is injured, the platelets in the area of the injury begin to clump together and stick to the edges of the cut. They also release messengers into the blood that perform a variety of functions: constricting the blood vessels to reduce bleeding, attracting more platelets to the area to enlarge the platelet plug, and initiating the work of plasma-based clotting factors, such as fibrinogen. Through a complex mechanism involving many steps and many clotting factors, the plasma protein fibrinogen is transformed into long, sticky threads of fibrin. Together, the platelets and the fibrin create an intertwined meshwork that forms a stable clot. This self-sealing aspect of the blood is crucial to survival.

IV -PRODUCTION AND ELIMINATION OF BLOOD CELLS

Blood is produced in the bone marrow, a tissue in the central cavity inside almost all of the bones in the body. In infants, the marrow in most of the bones is actively involved in blood cell formation. By later adult life, active blood cell formation gradually ceases in the bones of the arms and legs and concentrates in the skull, spine, ribs, and pelvis.

Red blood cells, white blood cells, and platelets grow from a single precursor cell, known as a hematopoietic stem cell. Remarkably, experiments have suggested that as few as 10 stem cells can, in four weeks, multiply into 30 trillion red blood cells, 30 billion white blood cells, and 1.2 trillion platelets—enough to replace every blood cell in the body.

Red blood cells have the longest average life span of any of the cellular elements of blood. A red blood cell lives 100 to 120 days after being released from the marrow into the blood. Over that period of time, red blood cells gradually age. Spent cells are removed by the spleen and, to a lesser extent, by the liver. The spleen and the liver also remove any red blood cells that become damaged, regardless of their age. The body efficiently recycles many components of the damaged cells, including parts of the hemoglobin molecule, especially the iron contained within it.

The majority of white blood cells have a relatively short life span. They may survive only 18 to 36 hours after being released from the marrow. However, some of the white blood cells are responsible for maintaining what is called immunologic memory. These memory cells retain knowledge of what infectious organisms the body has previously been exposed to. If one of those organisms returns, the memory cells initiate an extremely rapid response designed to kill the foreign invader. Memory cells may live for years or even decades before dying.

Memory cells make immunizations possible. An immunization, also called a vaccination or an inoculation, is a method of using a vaccine to make the human body immune to certain diseases. A vaccine consists of an infectious agent that has been weakened or killed in the laboratory so that it cannot produce disease when injected into a person, but can spark the immune system to generate memory cells and antibodies specific for the infectious agent. If the infectious agent should ever invade that vaccinated person in the future, these memory cells will direct the cells of the immune system to target the invader before it has the opportunity to cause harm.

Platelets have a life span of seven to ten days in the blood. They either participate in clot formation during that time or, when they have reached the end of their lifetime, are eliminated by the spleen and, to a lesser extent, by the liver.

V -BLOOD DISEASES

Many diseases are caused by abnormalities in the blood. These diseases are categorized by which component of the blood is affected.

A -Red Blood Cell Diseases

One of the most common blood diseases worldwide is anemia, which is characterized by an abnormally low number of red blood cells or low levels of hemoglobin. One of the major symptoms of anemia is fatigue, due to the failure of the blood to carry enough oxygen to all of the tissues.

The most common type of anemia, iron-deficiency anemia, occurs because the marrow fails to produce sufficient red blood cells. When insufficient iron is available to the bone marrow, it slows down its production of hemoglobin and red blood cells. In the United States, iron deficiency occurs most commonly due to poor nutrition. In other areas of the world, however, the most common causes of iron-deficiency anemia are certain infections that result in gastrointestinal blood loss and the consequent chronic loss of iron. Adding supplemental iron to the diet is often sufficient to cure iron-deficiency anemia.

Some anemias are the result of increased destruction of red blood cells, as in the case of sickle-cell anemia, a genetic disease most common in persons of African ancestry. The red blood cells of sickle-cell patients assume an unusual crescent shape, causing them to become trapped in some blood vessels, blocking the flow of other blood cells to tissues and depriving them of oxygen.

B -White Blood Cell Diseases

Some white blood cell diseases are characterized by an insufficient number of white blood cells. This can be caused by the failure of the bone marrow to produce adequate numbers of normal white blood cells, or by diseases that lead to the destruction of crucial white blood cells. These conditions result in severe immune deficiencies characterized by recurrent infections.

Any disease in which excess white blood cells are produced, particularly immature white blood cells, is called leukemia, or blood cancer. Many cases of leukemia are linked to gene abnormalities, resulting in unchecked growth of immature white blood cells. If this growth is not halted, it often results in the death of the patient. These genetic abnormalities are not inherited in the vast majority of cases, but rather occur after birth. Although some causes of these abnormalities are known, for example exposure to high doses of radiation or the chemical benzene, most remain poorly understood.

Treatment for leukemia typically involves the use of chemotherapy, in which strong drugs are used to target and kill leukemic cells, permitting normal cells to regenerate. In some cases, bone marrow transplants are effective. Much progress has been made over the last 30 years in the treatment of this disease. In one type of childhood leukemia, more than 80 percent of patients can now be cured of their disease.

C -Coagulation Diseases

One disease of the coagulation system is hemophilia, a genetic bleeding disorder in which one of the plasma clotting factors, usually factor VIII, is produced in abnormally low quantities, resulting in uncontrolled bleeding from minor injuries. Although individuals with hemophilia are able to form a good initial platelet plug when blood vessels are damaged, they are not easily able to form the meshwork that holds the clot firmly intact.

As a result, bleeding may occur some time after the initial traumatic event. Treatment for hemophilia relies on giving transfusions of factor VIII. Factor VIII can be isolated from the blood of normal blood donors but it also can be manufactured in a laboratory through a process known as gene cloning.

VI -BLOOD BANKS

The Red Cross and a number of other organizations run programs, known as blood banks, to collect, store, and distribute blood and blood products for transfusions. When blood is donated, its blood type is determined so that only appropriately matched blood is given to patients needing a transfusion. Before using the blood, the blood bank also tests it for the presence of disease-causing organisms, such as hepatitis viruses and human immunodeficiency virus (HIV), the cause of acquired immunodeficiency syndrome (AIDS).

This blood screening dramatically reduces, but does not fully eliminate, the risk to the recipient of acquiring a disease through a blood transfusion. Blood donation, which is extremely safe, generally involves giving about 400 to 500 ml (about 1 pt) of blood, which is only about 7 percent of a person's total blood.

VII -BLOOD IN NONHUMANS

One-celled organisms have no need for blood. They are able to absorb nutrients, expel wastes, and exchange gases with their environment directly. Simple multicelled marine animals, such as sponges, jellyfishes, and anemones, also do not have blood. They use the seawater that bathes their cells to perform the functions of blood. However, all more complex multicellular animals have some form of a circulatory system using blood. In some invertebrates, there are no cells analogous to red blood cells. Instead, hemoglobin, or the related copper compound heocyanin, circulates dissolved in the plasma.

The blood of complex multicellular animals tends to be similar to human blood, but there are also some significant differences, typically at the cellular level. For example, fish, amphibians, and reptiles possess red blood cells that have a nucleus, unlike the red blood cells of mammals. The immune system of invertebrates is more primitive than that of vertebrates, lacking the functionality associated with the white blood cell and antibody system found in mammals. Some arctic fish species produce proteins in their blood that act as a type of antifreeze, enabling them to survive in environments where the blood of other animals would freeze. Nonetheless, the essential transportation, communication, and protection functions that make blood essential to the continuation of life occur throughout much of the animal kingdom.

Environmental Effects of the Fossil Fuel Age

Over the last two centuries, human activity has transformed the chemistry of Earth's water and air, altered the face of Earth itself, and rewoven the web of life. Why has this time period, more than any other, brought so much widespread environmental change? The reasons are many and complex. But a major influence surely is the use of fossil fuels, which has made far more energy available to more people than had ever been available before.

By 1990, humans were using about 80 times as much energy as was being used in 1800. The great majority of this energy was derived from fossil fuels. The availability and use of this new energy source has allowed people to produce more and consume more. Indirectly, this energy source caused a rapid increase in population as people developed much more efficient means of agriculture—such as mechanized farming—that required the use of fossil fuels. Improved farming techniques brought about an increase in food supply, which fostered the population growth. By the end of the 1990s, the human population was about six times what it was in 1800. Widespread changes to the environment resulted from other factors as well. The breakneck pace of urbanization is a factor, as is the equally dizzying speed of technological change. No less important a factor in environmental change is the heightened emphasis of modern governments on economic growth. All of these trends are interrelated, each one helping to advance the others.

Together, they have shaped the evolution of human society in modern times. These growth trends have recast the relationships between humanity and other inhabitants of Earth.

For hundreds of thousands of years, human beings and their predecessors have both deliberately and accidentally altered their environments. But only recently, with the harnessing of fossil fuels, has humankind acquired the power to effect thorough changes on air, water, soils, plants, and animals. Armed with fossil fuels, people have changed the environment in ways they never had in pre-modern times—for example, devastating natural habitats and wildlife with oil spills. People have also been able to bring about environmental change much more rapidly, through acceleration of old activities such as deforestation.

Origins of Fossil Fuels

Fossil fuels include coal, natural gas, and petroleum (also known as oil or crude oil), which are the petrified and liquefied remains of millions of years' accumulation of decayed plant life. When fossil fuels are burned, their chemical energy becomes heat energy, which, by means of machines such as engines and turbines, is converted into mechanical or electrical energy.

Coal first became an important industrial fuel during the 11th and 12th centuries in China, where iron manufacturing consumed great quantities of this resource. The first major usage of coal as a domestic fuel began in 16th-century London, England. During the Industrial Revolution, which began in the 18th century, coal became a key fuel for industry, powering most steam engines.

Coal was the primary fossil fuel until the middle of the 20th century, when oil replaced it as the fuel of choice in industry, transportation, and other fields. Deep drilling for petroleum was pioneered in western Pennsylvania in 1859, and the first large oil fields were tapped in southeastern Texas in 1901. The world's biggest oil fields were accessed in the 1940s in Saudi Arabia and in the 1960s in Siberia. Why did oil overshadow coal as the fuel of choice? Oil has certain advantages over coal. It is more efficient than coal, providing more energy per unit of weight than coal does. Oil also causes less pollution and works better in small engines. Oil is less plentiful than coal, however. When the world runs low on oil, copious supplies of coal will remain available.

Modern Air Pollution

The outermost layer of the Earth's living environment is the atmosphere, a mixture of gases surrounding the planet. The atmosphere contains a thin layer called ozone, which protects all life on Earth from harmful ultraviolet radiation from the Sun. For most of human history, people had very little effect on the atmosphere. For many thousands of years, humans routinely burned vegetation, causing some intermittent air pollution. In ancient times, the smelting of ores, such as copper ore, released metals that traveled in the atmosphere from the shores of the Mediterranean Sea as far as Greenland. With the development of fossil fuels, however, much more intense air pollution began to trouble humanity.

Before widespread use of fossil fuels, air pollution typically affected cities more than it did rural areas because of the concentration of combustion in cities. People in cold-climate urban areas kept warm by burning wood, but local wood supplies were soon exhausted. As a result of the limited supply, wood became expensive. People then burned comparatively little amounts of wood and heated their homes less. The first city to resolve this problem was London, where residents began using coal to heat their buildings. By the 1800s, half a million chimneys were releasing coal smoke, soot, ash, and sulfur dioxide into the London air.

The development of steam engines in the 18th century introduced coal to industry. The resultant growth from the Industrial Revolution meant more steam engines, more factory chimneys, and, thus, more air pollution. Skies darkened in the industrial heartlands of Britain, Belgium, Germany, and the United States. Cities that combined energy-intensive industries such as iron and steel manufacturing, and coal-heated buildings, were routinely shrouded in smoke and bathed in sulfur dioxide. Pittsburgh, Pennsylvania, one of the United States' major industrial cities at the time, was sometimes referred to as "Hell with the lid taken off." The coal consumption of some industries was so great that it could pollute the skies over entire regions, as was the case in the Ruhr region in Germany and around Hanshin, the area near Ōsaka, Japan.

Early Air Pollution Control

Efforts at smoke abatement were largely ineffective until about 1940, so residents of industrial cities and regions suffered the consequences of life with polluted air. During the Victorian Age in England, dusting household surfaces twice a day to keep up with the dustfall was not uncommon. Residents of industrial cities witnessed the loss of pine trees and some wildlife, due to the high levels of sulfur dioxide. These people suffered rates of pneumonia and bronchitis far higher than those of their ancestors, their relatives living elsewhere, or their descendants.

After 1940, leaders of industrial cities and regions managed to reduce the severity of coal-based air pollution. St. Louis, Missouri, was the first city in the world to make smoke abatement a high priority. Pittsburgh and other U.S. cities followed during the late 1940s and 1950s. London took effective steps during the mid-1950s after the killer fog, an acute bout of pollution in December of 1952, took some 4,000 lives. Germany and Japan made strides toward smoke abatement during the 1960s, using a combination of taller smokestacks, smokestack filters and scrubbers, and the substitution of other fuels for coal.

Even as smoke abatement continued, however, cities acquired new and more complex air pollution problems. As cars became commonplace—first in the United States during the 1920s and then in Western Europe and Japan during the 1950s and 1960s—tailpipe emissions added to the air pollution already flowing out of chimneys and smokestacks. Auto exhaust contained different kinds of pollutants, such as carbon monoxide, nitrous oxide, and lead. Therefore cars, together with new industries, such as the petrochemical industry, complicated and intensified the world's air pollution problems. Photochemical smog, which is caused by sunlight's impact on elements of auto exhaust, became a serious health menace in cities where abundant sunshine combined with frequent temperature change. The world's worst smog was brewed in sunny, car-clogged cities, such as Athens, Greece; Bangkok, Thailand; Mexico City, Mexico; and Los Angeles, California.

In addition to these local and regional pollution problems, during the late 20th century human activity began to take its toll on the atmosphere. The increased carbon dioxide levels in the atmosphere after 1850, which were mainly a consequence of burning fossil fuels, raised the efficiency with which the air retains the sun's heat. This greater heat retention brought the threat of global warming, an overall increase in Earth’s temperature. Yet another threat to the atmosphere was caused by chemicals known as chlorofluorocarbons, which were invented in 1930 and used widely in industry and as refrigerants after 1950. When chlorofluorocarbons float up to the stratosphere (the upper layer of Earth’s atmosphere), they cause the ozone layer to become thinner, hampering its ability to block harmful ultraviolet radiation.

Water Pollution

Water has always been a vital resource for human beings—at first just for drinking, later for washing, and eventually for irrigation. With the power conferred by fossil fuels and modern technology, people have rerouted rivers, pumped up deep groundwater, and polluted the Earth’s water supply as never before.

Irrigation, though an ancient practice, affected only small parts of the world until recently. During the 1800s, irrigation practices spread quickly, driven by advances in engineering and increased demand for food by the world’s growing population. In India and North America, huge networks of dams and canals were built. The 1900s saw the construction of still larger dams in these countries, as well as in Central Asia, China, and elsewhere. After the 1930s, dams built for irrigation also served to generate hydroelectric power. Between 1945 and 1980, most of the world's rivers that had met engineers’ criteria for suitability had acquired dams.

Because they provided electric power as well as irrigation water, dams made life easier for millions of people. Convenience came at a price, however, as dams changed established water ecosystems that had developed over the course of centuries. In the Columbia River in western North America, for example, salmon populations suffered because dams blocked the annual migrations of the salmon. In Egypt, where a large dam spanned the Nile at Aswan after 1971, many humans and animals paid the price. Mediterranean sardines died and the fisherman who caught these fish lost their business. Farmers had to resort to chemical fertilizers because the dam prevented the Nile’s spring flooding and the resultant annual coating of fertile silt on land along the river. In addition, many Egyptians who drank Nile water, which carried increasing amounts of fertilizer runoff, experienced negative health effects. In Central Asia, the Aral Sea paid the price. After 1960 this sea shrank because the waters that fed into it were diverted to irrigate cotton fields.

River water alone did not suffice to meet the water needs of agriculture and cities. Groundwater in many parts of the world became an essential source of water. This source was available at low cost, because fossil fuels made pumping much easier. For example, after 1930 an economy based on grain and livestock emerged on the High Plains, from Texas to the Dakotas. This economy drew water from the Ogallala Aquifer, a vast underground reservoir. To meet the drinking, washing, and industrial needs of their growing populations, cities such as Barcelona, Spain; Beijing, China; and Mexico City, pumped up groundwater. Beijing and Mexico City began sinking slowly into the ground as they pumped out much of their underlying water. As groundwater supplies dwindled, both cities found they needed to bring water in from great distances. By 1999 humanity was using about 20 times as much fresh water as was used in 1800.

Not only was the water use increasing, but more of it was becoming polluted by human use. While water pollution had long existed in river water that flowed through cities, such as the Seine in Paris, France, the fossil fuel age changed the scope and character of water pollution. Water usage increased throughout this era, and a far wider variety of pollutants contaminated the world's water supplies. For most of human history, water pollution was largely biological, caused mainly by human and animal wastes. However, industrialization introduced countless chemicals into the waters of the world, complicating pollution problems.

Efforts to Control Water Pollution

Until the early 20th century, biological pollution of the world's lakes and rivers remained a baffling problem. Then experiments in filtration and chemical treatment of water proved fruitful. In Europe and North America, sewage treatment and water filtration assured a cleaner and healthier water supply. As late as the 1880s in Chicago, Illinois, thousands of people died each year from waterborne diseases, such as typhoid fever. By 1920, though, Chicago's water no longer carried fatal illnesses. Many communities around the world, especially in poor countries such as India and Nigeria, could not afford to invest in sewage treatment and water filtration plants, however.

As was the case with air pollution, the industrialization and technological advances of the 20th century brought increasing varieties of water pollution. Scientists invented new chemicals that did not exist in nature, and a few of these chemicals turned out to be very useful in manufacturing and in agriculture. Unfortunately, a few of these also turned out to be harmful pollutants. After 1960 chemicals called polychlorinated biphenyls (PCBs) turned up in dangerous quantities in North American waters, killing and damaging aquatic life and the creatures that eat these plants and animals. After 1970, legislation in North America and Europe substantially reduced point pollution, or water pollution derived from single sources. But nonpoint pollution, such as pesticide-laced runoff from farms, proved much harder to control. The worst water pollution prevailed in poorer countries where biological pollution continued unabated, while chemical pollution from industry or agriculture emerged to complement the biological pollution. In the late 1900s, China probably suffered the most from the widest variety of water pollution problems.

Soil Pollution

During the era of fossil fuels, the surface of Earth also has undergone remarkable change. The same substances that have polluted the air and water often lodge in the soil, occasionally in dangerous concentrations that threaten human health. While this situation normally happened only in the vicinity of industries that generated toxic wastes, the problem of salinization, or salting, which was associated with irrigation, was more widespread.

Although irrigation has always brought the risk of destroying soils by waterlogging and salinization—the ancient middle-eastern civilization of Mesopotamia probably undermined its agricultural base this way—the modern scale of irrigation has intensified this problem around the world. By the 1990s, fields ruined by salinization were being abandoned as fast as engineers could irrigate new fields. Salinization has been the most severe in dry lands where evaporation occurs the fastest, such as in Mexico, Australia, Central Asia, and the Southwestern United States.

Soil erosion due to human activity was a problem long before salinization was. Modern soil erosion diminished the productivity of agriculture. This problem was worst in the 1800s in the frontier lands newly opened to pioneer settlement in the United States, Canada, Australia, New Zealand, Argentina, and elsewhere. Grasslands that had never been plowed before became vulnerable to wind erosion, which reached disastrous proportions during droughts, such as those during the 1930s in the Dust Bowl of Kansas and Oklahoma. The last major clearing of virgin grassland took place in the Union of Soviet Socialist Republics (USSR) in the 1950s, when Premier Nikita Khrushchev decided to convert northern Kazakhstan into a wheat belt. Fossil fuels also played a crucial role at this time, because railroads and steamships carried the grain and beef raised in these frontiers to distant markets.

By the late 20th century, pioneer settlement had shifted away from the world's grasslands into tropical and mountain forest regions. After 1950, farmers in Asia, Africa, and Latin America increasingly sought land in little-cultivated forests. Often these forests, such as those in Central America or the Philippines, were mountainous and subject to heavy rains. In order to cultivate this land, farmers deforested these mountainsides, which exposed them to heavy rains and invited soil erosion. Erosion caused in this manner stripped soils in the Andes of Bolivia, in the Himalayas of Nepal and northern India, and in the rugged terrain of Rwanda and Burundi. Depleted soils made life harder for farmers in these and other lands.

The impact of soil erosion does not stop with the loss of soil. Eroded soil does not simply disappear. Rather, it flows downhill and downstream, only to rest somewhere else. Often this soil has lodged in inconvenient places, silting up dam reservoirs or covering roads. Within only a few years of being built, some dams in Algeria and China became useless because they were clogged by soil erosion originating upstream.

Animal and Plant Life

Human activity has affected the world's plants and animals no less than it has the air, water, and soil. For millions of years, life evolved without much impact from human beings. However, as early as the first settlements of Australia and North America, human beings probably caused mass extinctions, either through hunting or through the use of fire. With the domestication of animals, which began perhaps 10,000 years ago, humanity came to play a more active role in biological evolution. By the 1800s and 1900s, the role that human beings played in species survival had expanded to the extent that many species survive only because human beings allow it.

Some animal species survive in great numbers thanks to us. For example, today there are about 10 billion chickens on Earth—about thirteen to fifteen times as many as there were a century ago. This is because people like to eat chickens, so they are raised for this purpose. Similarly, we protect cattle, sheep, goats, and a few other domesticated animals in order to make use of them. Inadvertently, modern civilizations have ensured the survival of certain other animals. Rat populations propagate because of all of the food available to them, since humans store so much food and generate so much garbage. Squirrels prosper in large part because we have created suburban landscapes with few predators.

Even as modern human beings intentionally or unintentionally encourage the survival of a few species, humans threaten many more. Modern technology and fuels have made hunting vastly more efficient, bringing animals such as the blue whale and the North American bison to the edge of extinction. Many other animals, most notably tropical forest species, suffer from destruction of their preferred habitats. Quite inadvertently, and almost unconsciously, humankind has assumed a central role in determining the fate of many species and the health of Earth's water, air, and soil. Humans have therefore assumed a central role in biological evolution.

The environmental history of the last two centuries has been one of enormous change. In a mere 200 years, humanity has altered Earth more drastically than since the dawn of agriculture about 10,000 years ago. Our vital air, water, and soil have been jeopardized; the very web of life hangs on our whims. For the most part, human beings have never been more successful nor led easier lives. The age of fossil fuels is changing the human condition in ways previously unimaginable. But whether we understand the impact—and are willing to accept it—remains an unanswered question.
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Darwin, Charles Robert

I –INTRODUCTION

Darwin, Charles Robert (1809-1882), British scientist, who laid the foundation of modern evolutionary theory with his concept of the development of all forms of life through the slow-working process of natural selection. His work was of major influence on the life and earth sciences and on modern thought in general.

Born in Shrewsbury, Shropshire, England, on February 12, 1809, Darwin was the fifth child of a wealthy and sophisticated English family. His maternal grandfather was the successful china and pottery entrepreneur Josiah Wedgwood; his paternal grandfather was the well-known 18th-century physician and savant Erasmus Darwin. After graduating from the elite school at Shrewsbury in 1825, young Darwin went to the University of Edinburgh to study medicine. In 1827 he dropped out of medical school and entered the University of Cambridge, in preparation for becoming a clergyman of the Church of England. There he met two stellar figures: Adam Sedgwick, a geologist, and John Stevens Henslow, a naturalist. Henslow not only helped build Darwin's self-confidence but also taught his student to be a meticulous and painstaking observer of natural phenomena and collector of specimens. After graduating from Cambridge in 1831, the 22-year-old Darwin was taken aboard the English survey ship HMS Beagle, largely on Henslow's recommendation, as an unpaid naturalist on a scientific expedition around the world.

II -VOYAGE OF THE BEAGLE

Darwin's job as naturalist aboard the Beagle gave him the opportunity to observe the various geological formations found on different continents and islands along the way, as well as a huge variety of fossils and living organisms. In his geological observations, Darwin was most impressed with the effect that natural forces had on shaping the earth's surface.

At the time, most geologists adhered to the so-called catastrophist theory that the earth had experienced a succession of creations of animal and plant life, and that each creation had been destroyed by a sudden catastrophe, such as an upheaval or convulsion of the earth's surface (see Geology: History of Geology: Geology in the 18th and 19th Centuries). According to this theory, the most recent catastrophe, Noah's flood, wiped away all life except those forms taken into the ark. The rest were visible only in the form of fossils. In the view of the catastrophists, species were individually created and immutable, that is, unchangeable for all time.

The catastrophist viewpoint (but not the immutability of species) was challenged by the English geologist Sir Charles Lyell in his three-volume work Principles of Geology (1830-1833). Lyell maintained that the earth's surface is undergoing constant change, the result of natural forces operating uniformly over long periods.

Aboard the Beagle, Darwin found himself fitting many of his observations into Lyell's general uniformitarian view. Beyond that, however, he realized that some of his own observations of fossils and living plants and animals cast doubt on the Lyell-supported view that species were specially created. He noted, for example, that certain fossils of supposedly extinct species closely resembled living species in the same geographical area. In the Galápagos Islands, off the coast of Ecuador, he also observed that each island supported its own form of tortoise, mockingbird, and finch; the various forms were closely related but differed in structure and eating habits from island to island. Both observations raised the question, for Darwin, of possible links between distinct but similar species.

III -THEORY OF NATURAL SELECTION

After returning to England in 1836, Darwin began recording his ideas about changeability of species in his Notebooks on the Transmutation of Species. Darwin's explanation for how organisms evolved was brought into sharp focus after he read An Essay on the Principle of Population (1798), by the British economist Thomas Robert Malthus, who explained how human populations remain in balance. Malthus argued that any increase in the availability of food for basic human survival could not match the geometrical rate of population growth. The latter, therefore, had to be checked by natural limitations such as famine and disease, or by social actions such as war.

Darwin immediately applied Malthus's argument to animals and plants, and by 1838 he had arrived at a sketch of a theory of evolution through natural selection (see Species and Speciation). For the next two decades he worked on his theory and other natural history projects. (Darwin was independently wealthy and never had to earn an income.) In 1839 he married his first cousin, Emma Wedgwood, and soon after, moved to a small estate, Down House, outside London. There he and his wife had ten children, three of whom died in infancy.

Darwin's theory was first announced in 1858 in a paper presented at the same time as one by Alfred Russel Wallace, a young naturalist who had come independently to the theory of natural selection. Darwin's complete theory was published in 1859, in On the Origin of Species. Often referred to as the "book that shook the world," the Origin sold out on the first day of publication and subsequently went through six editions.

Darwin's theory of evolution by natural selection is essentially that, because of the food-supply problem described by Malthus, the young born to any species intensely compete for survival. Those young that survive to produce the next generation tend to embody favorable natural variations (however slight the advantage may be)—the process of natural selection—and these variations are passed on by heredity. Therefore, each generation will improve adaptively over the preceding generations, and this gradual and continuous process is the source of the evolution of species. Natural selection is only part of Darwin's vast conceptual scheme; he also introduced the concept that all related organisms are descended from common ancestors. Moreover, he provided additional support for the older concept that the earth itself is not static but evolving.

IV -REACTIONS TO THE THEORY

The reaction to the Origin was immediate. Some biologists argued that Darwin could not prove his hypothesis. Others criticized Darwin's concept of variation, arguing that he could explain neither the origin of variations nor how they were passed to succeeding generations. This particular scientific objection was not answered until the birth of modern genetics in the early 20th century (see Heredity; Mendel's Laws). In fact, many scientists continued to express doubts for the following 50 to 80 years. The most publicized attacks on Darwin's ideas, however, came not from scientists but from religious opponents. The thought that living things had evolved by natural processes denied the special creation of humankind and seemed to place humanity on a plane with the animals; both of these ideas were serious contradictions to orthodox theological opinion.

V -LATER YEARS

Darwin spent the rest of his life expanding on different aspects of problems raised in the Origin. His later books—including The Variation of Animals and Plants Under Domestication (1868), The Descent of Man (1871), and The Expression of the Emotions in Man and Animals (1872)—were detailed expositions of topics that had been confined to small sections of the Origin. The importance of his work was well recognized by his contemporaries; Darwin was elected to the Royal Society (1839) and the French Academy of Sciences (1878). He was also honored by burial in Westminster Abbey after he died in Downe, Kent, on April 19, 1882.

Adaptation

I –INTRODUCTION

Adaptation word used by biologists in two different senses, both of which imply the accommodation of a living organism to its environment. One form of adaptation, called physiological adaptation, involves the acclimatization of an individual organism to a sudden change in environment. The other kind of adaptation, discussed here, occurs during the slow course of evolution and hence is called evolutionary adaptation.

II -MECHANISMS OF ADAPTATION

Evolutionary adaptations are the result of the competition among individuals of a particular species over many generations in response to an ever-changing environment, including other animals and plants. Certain traits are culled by natural selection (see Evolution), favoring those individual organisms that produce the most offspring. This is such a broad concept that, theoretically, all the features of any animal or plant could be considered adaptive. For example, the leaves, trunk, and roots of a tree all arose by selection and help the individual tree in its competition for space, soil, and sunlight.

Biologists have been accused of assuming adaptive ness for all such features of a species, but few cases have actually been demonstrated. Indeed, biologists find it difficult to be certain whether any particular structure of an organism arose by selection and hence can be called adaptive or whether it arose by chance and is selectively neutral.

The best example of an evolutionary development with evidence for adaptation is mimicry. Biologists can show experimentally that some organisms escape predators by trying to be inconspicuous and blend into their environment and that other organisms imitate the coloration of species distasteful to predators. These tested cases are only a handful, however, and many supposed cases of adaptation are simply assumed.

On the contrary, it is possible that some features of an organism may be retained because they are adaptive for special, limited reasons, even though they may be maladaptive on the whole. The large antlers of an elk or moose, for example, may be effective in sexual selection for mating but could well be maladaptive at all other times of the year. In addition, a species feature that now has one adaptive significance may have been produced as an adaptation to quite different circumstances. For example, lungs probably evolved in adaptation to life in water that sometimes ran low on oxygen. Fish with lungs were then “preadapted” in a way that accidentally allowed their descendants to become terrestrial.

III -ADAPTIVE RADIATION

Because the environment exerts such control over the adaptations that arise by natural selection—including the coadaptations of different species evolving together, such as flowers and pollinators—the kind of organism that would fill a particular environmental niche ought to be predictable in general terms. An example of this process of adaptative radiation, or filling out of environmental niches by the development of new species, is provided by Australia.

When Australia became a separate continent some 60 million years ago, only monotremes and marsupials lived there, with no competition from the placental mammals that were emerging on other continents. Although only two living monotremes are found in Australia today, the marsupials have filled most of the niches open to terrestrial mammals on that continent. Because Australian habitats resemble those in other parts of the world, marsupial equivalents can be found to the major placental herbivores, carnivores, and even rodents and moles.

This pattern can be observed on a restricted scale as well. In some sparsely populated islands, for example, one species of bird might enter the region, find little or no competition, and evolve rapidly into a number of species adapted to the available niches. A well-known instance of such adaptive radiation was discovered by Charles Darwin in the Galápagos Islands. He presumed, probably correctly, that one species of finch colonized the islands thousands of years ago and gave rise to the 14 species of finchlike birds that exist there now. Thus, one finch behaves like a warbler, another like a woodpecker, and so on. The greatest differences in their appearance lie in the shapes of the bills, adapted to the types of food each species eats.

IV -ANALOGY AND HOMOLOGY

When different species are compared, some adaptive features can be described as analogous or homologous. For example, flight requires certain rigid aeronautical principles of design; yet birds, bats, and insects have all conquered the air. In this case the flight structures are said to be analogous—that is, they have different embryological origins but perform the same function. By contrast, structures that arise from the same structures in the embryo but are used in entirely different kinds of functions, such as the forelimb of a rat and the wing of a bat, are said to be homologous.

Plate Tectonics

I –INTRODUCTION

Plate Tectonics, theory that the outer shell of the earth is made up of thin, rigid plates that move relative to each other. The theory of plate tectonics was formulated during the early 1960s, and it revolutionized the field of geology. Scientists have successfully used it to explain many geological events, such as earthquakes and volcanic eruptions as well as mountain building and the formation of the oceans and continents.

Plate tectonics arose from an earlier theory proposed by German scientist Alfred Wegener in 1912. Looking at the shapes of the continents, Wegener found that they fit together like a jigsaw puzzle. Using this observation, along with geological evidence he found on different continents, he developed the theory of continental drift, which states that today's continents were once joined together into one large landmass.

Geologists of the 1950s and 1960s found evidence supporting the idea of tectonic plates and their movement. They applied Wegener's theory to various aspects of the changing earth and used this evidence to confirm continental drift. By 1968 scientists integrated most geologic activities into a theory called the New Global Tectonics, or more commonly, Plate Tectonics.

II -TECTONIC PLATES

Tectonic plates are made of either oceanic or continental crust and the very top part of the mantle, a layer of rock inside the earth. This crust and upper mantle form what is called the lithosphere. Under the lithosphere lies a fluid rock layer called the asthenosphere. The rocks in the asthenosphere move in a fluid manner because of the high temperatures and pressures found there. Tectonic plates are able to float upon the fluid asthenosphere because they are made of rigid lithosphere. See also Earth: Plate Tectonics.

A -Continental Crust

The earth's solid surface is about 40 percent continental crust. Continental crust is much older, thicker and less dense than oceanic crust. The thinnest continental crust, between plates that are moving apart, is about 15 km (about 9 mi) thick. In other places, such as mountain ranges, the crust may be as much as 75 km (47 mi) thick. Near the surface, it is composed of rocks that are felsic (made up of minerals including feldspar and silica). Deeper in the continental crust, the composition is mafic (made of magnesium, iron, and other minerals).

B -Oceanic Crust

Oceanic crust makes up the other 60 percent of the earth's solid surface. Oceanic crust is, in general, thin and dense. It is constantly being produced at the bottom of the oceans in places called mid-ocean ridges—undersea volcanic mountain chains formed at plate boundaries where there is a build-up of ocean crust. This production of crust does not increase the physical size of the earth, so the material produced at mid-ocean ridges must be recycled, or consumed, somewhere else. Geologists believe it is recycled back into the earth in areas called subduction zones, where one plate sinks underneath another and the crust of the sinking plate melts back down into the earth. Oceanic crust is continually recycled so that its age is generally not greater than 200 million years. Oceanic crust averages between 5 and 10 km (between 3 and 6 mi) thick. It is composed of a top layer of sediment, a middle layer of rock called basalt, and a bottom layer of rock called gabbro. Both basalt and gabbro are dark-colored igneous, or volcanic, rocks.

C -Plate Sizes

Currently, there are seven large and several small plates. The largest plates include the Pacific plate, the North American plate, the Eurasian plate, the Antarctic plate, and the African plate. Smaller plates include the Cocos plate, the Nazca plate, the Caribbean plate, and the Gorda plate. Plate sizes vary a great deal. The Cocos plate is 2000 km (1400 mi) wide, while the Pacific plate is the largest plate at nearly 14,000 km (nearly 9000 mi) wide.

III -PLATE MOVEMENT

Geologists study how tectonic plates move relative to a fixed spot in the earth's mantle and how they move relative to each other. The first type of motion is called absolute motion, and it can lead to strings of volcanoes. The second kind of motion, called relative motion, leads to different types of boundaries between plates: plates moving apart from one another form a divergent boundary, plates moving toward one another form a convergent boundary, and plates that slide along one another form a transform plate boundary. In rare instances, three plates may meet in one place, forming a triple junction. Current plate movement is making the Pacific Ocean smaller, the Atlantic Ocean larger, and the Himalayan mountains taller.

A -Measuring Plate Movement

Geologists discovered absolute plate motion when they found chains of extinct submarine volcanoes. A chain of dead volcanoes forms as a plate moves over a plume, a source of magma, or molten rock, deep within the mantle. These plumes stay in one spot, and each one creates a hot spot in the plate above the plume. These hot spots can form into a volcano on the surface of the earth. An active volcano indicates a hot spot as well as the youngest region of a volcanic chain. As the plate moves, a new volcano forms in the plate over the place where the hot spot occurs. The volcanoes in the chain get progressively older and become extinct as they move away from the hot spot (see Hawaii: Formation of the Islands and Volcanoes).

Scientists use hot spots to measure the speed of tectonic plates relative to a fixed point. To do this, they determine the age of extinct volcanoes and their distance from a hot spot. They then use these numbers to calculate how far the plate has moved in the time since each volcano formed. Today, the plates move at velocities up to 18.5 cm per year (7.3 in per year). On average, they move nearly 4 to 7 cm per year (2 to 3 in per year).

B -Divergent Plate Boundaries

Divergent plate boundaries occur where two plates are moving apart from each other. When plates break apart, the lithosphere thins and ruptures to form a divergent plate boundary. In the oceanic crust, this process is called seafloor spreading, because the splitting plates are spreading apart from each other. On land, divergent plate boundaries create rift valleys—deep valley depressions formed as the land slowly splits apart.

When seafloor spreading occurs, magma, or molten rock material, rises to the sea floor surface along the rupture. As the magma cools, it forms new oceanic crust and lithosphere. The new lithosphere is less dense, so it rises, or floats, higher above older lithosphere, producing long submarine mountain chains known as mid-ocean ridges. The Mid-Atlantic Ridge is an underwater mountain range created at a divergent plate boundary in the middle of the Atlantic Ocean. It is part of a worldwide system of ridges made by seafloor spreading. The Mid-Atlantic Ridge is currently spreading at a rate of 2.5 cm per year (1 in per year). The mid-ocean ridges today are 60,000 km (about 40,000 mi) long, forming the largest continuous mountain chain on earth. Earthquakes, faults, underwater volcanic eruptions, and vents, or openings, along the mountain crests produce rugged seafloor features, or topography.

Divergent boundaries on land cause rifting, in which broad areas of land are uplifted, or moved upward. These uplifts and faulting along the rift result in rift valleys. Examples of rift valleys are found at the Krafla Volcano rift area in Iceland as well as at the East African Rift Zone—part of the Great Rift Valley that extends from Syria to Mozambique and out to the Red Sea. In these areas, volcanic eruptions and shallow earthquakes are common.

C -Convergent Plate Boundaries

Convergent plate boundaries occur where plates are consumed, or recycled back into the earth's mantle. There are three types of convergent plate boundaries: between two oceanic plates, between an oceanic plate and a continental plate, and between two continental plates. Subduction zones are convergent regions where oceanic crust is thrust below either oceanic crust or continental crust. Many earthquakes occur at subduction zones, and volcanic ridges and oceanic trenches form in these areas.

In the ocean, convergent plate boundaries occur where an oceanic plate descends beneath another oceanic plate. Chains of active volcanoes develop 100 to 150 km (60 to 90 mi) above the descending slab as magma rises from under the plate. Also, where the crust slides down into the earth, a trench forms. Together, the volcanoes and trench form an intra-oceanic island arc and trench system. A good example of such a system is the Mariana Trench system in the western Pacific Ocean, where the Pacific plate is descending under the Philippine plate. In these areas, earthquakes are frequent but not large. Stress in and behind the arc often causes the arc and trench system to move toward the incoming plate, which opens small ocean basins behind the arc. This process is called back-arc seafloor spreading.

Convergent boundaries that occur between the ocean and land create continental margin arc and trench systems near the margins, or edges, of continents. Volcanoes also form here. Stress can develop in these areas and cause the rock layers to fold, leading to earthquake faults, or breaks in the earth's crust called thrust faults. The folding and thrust faulting thicken the continental crust, producing high mountains. Many of the world's large destructive earthquakes and major mountain chains, such as the Andes Mountains of western South America, occur along these convergent plate boundaries.

When two continental plates converge, the incoming plate drives against and under the opposing continent. This often affects hundreds of miles of each continent and, at times, doubles the normal thickness of continental crust. Colliding continents cause earthquakes and form mountains and plateaus. The collision of India with Asia has produced the Himalayan Mountains and Tibetan Plateau.

D -Transform Plate Boundaries

A transform plate boundary, also known as a transform fault system, forms as plates slide past one another in opposite directions without converging or diverging. Early in the plate tectonic revolution, geologists proposed that transform faults were a new class of fault because they "transformed" plate motions from one plate boundary to another. Canadian geophysicist J. Tuzlo Wilson studied the direction of faulting along fracture zones that divide the mid-ocean ridge system and confirmed that transform plate boundaries were different than convergent and divergent boundaries. Within the ocean, transform faults are usually simple, straight fault lines that form at a right angle to ocean ridge spreading centers. As plates slide past each other, the transform faults can divide the centers of ocean ridge spreading. By cutting across the ridges of the undersea mountain chains, they create steep cliff slopes. Transform fault systems can also connect spreading centers to subduction zones or other transform fault systems within the continental crust. As a transform plate boundary cuts perpendicularly across the edges of the continental crust near the borders of the continental and oceanic crust, the result is a system such as the San Andreas transform fault system in California.

E -Triple Junction

Rarely, a group of three plates, or a combination of plates, faults, and trenches, meet at a point called a triple junction. The East African Rift Zone is a good example of a triple plate junction. The African plate is splitting into two plates and moving away from the Arabian plate as the Red Sea meets the Gulf of Aden. Another example is the Mendocino Triple Junction, which occurs at the intersection of two transform faults (the San Andreas and Mendocino faults) and the plate boundary between the Pacific and Gorda plates.

F -Current Plate Movemen

Plate movement is changing the sizes of our oceans and the shapes of our continents. The Pacific plate moves at an absolute motion rate of 9 cm per year (4 in per year) away from the East Pacific Rise spreading center, the undersea volcanic region in the eastern Pacific Ocean that runs parallel to the western coast of South America. On the other side of the Pacific Ocean, near Japan, the Pacific plate is being subducted, or consumed under, the oceanic arc systems found there. The Pacific Ocean is getting smaller as the North and South American plates move west. The Atlantic Ocean is getting larger as plate movement causes North and South America to move away from Europe and Africa. Since the Eurasian and Antarctic plates are nearly stationary, the Indian Ocean at present is not significantly expanding or shrinking. The plate that includes Australia is just beginning to collide with the plate that forms Southeast Asia, while India's plate is still colliding with Asia. India moves north at 5 cm per year (2 in per year) as it crashes into Asia, while Australia moves slightly farther away from Antarctica each year.

IV -CAUSES OF PLATE MOTION

Although plate tectonics has explained most of the surface features of the earth, the driving force of plate tectonics is still unclear. According to geologists, a model that explains plate movement should include three forces. Those three forces are the pull of gravity; convection currents, or the circulating movement of fluid rocky material in the mantle; and thermal plumes, or vertical columns of molten rocky material in the mantle.

A -Plate Movement Caused by Gravity

Geologists believe that tectonic plates move primarily as a result of their own weight, or the force of gravity acting on them. Since the plates are slightly denser than the underlying asthenosphere, they tend to sink. Their weight causes them to slide down gentle gradients, such as those formed by the higher ocean ridge crests, to the lower subduction zones. Once the plate's leading edge has entered a subduction zone and penetrated the mantle, the weight of the slab itself will tend to pull the rest of the plate toward the trench. This sinking action is known as slab-pull because the sinking plate edge pulls the remainder of the plate behind it. Another kind of action, called ridge-push, is the opposite of slab-pull, in that gravity also causes plates to slide away from mid-ocean ridges. Scientists believe that plates pushing against one another also causes plate movement.

B -Convection Currents

In 1929 British geologist Arthur Holmes proposed the concept of convection currents—the movement of molten material circulating deep within the earth—and the concept was modified to explain plate movement. A convection current occurs when hot, molten, rocky material floats up within the asthenosphere, then cools as it approaches the surface. As it cools, the material becomes denser and begins to sink again, moving in a circular pattern. Geologists once thought that convection currents were the primary driving force of plate movement. They now believe that convection currents are not the primary cause, but are an effect of sinking plates that contributes to the overall movement of the plates.

C -Thermal Plumes

Some scientists have proposed the concept of thermal plumes, vertical columns of molten material, as an additional force of plate movement. Thermal plumes do not circulate like convection currents. Rather, they are columns of material that rise up through the asthenosphere and appear on the surface of the earth as hot spots. Scientists estimate thermal plumes to be between 100 and 250 km (60 and 160 mi) in diameter. They may originate within the asthenosphere or even deeper within the earth at the boundary between the mantle and the core.

V -EXTRATERRESTRIAL PLATE TECTONICS

Scientists have also observed tectonic activity and fracturing on several moons of other planets in our solar system. Starting in 1985, images from the Voyager probes indicated that Saturn's satellite Enceladus and Uranus' moon Miranda also show signs of being tectonically active. In 1989 the Voyager probes sent photographs and data to Earth of volcanic activity on Neptune's satellite Triton. In 1995 the Galileo probe began to send data and images of tectonic activity on three of Jupiter's four Galilean satellites. The information that scientists gather from space missions such as these helps increase their understanding of the solar system and our planet. They can apply this knowledge to better understand the forces that created the earth and that continue to act upon it.

Scientists believe that Enceladus has a very tectonically active surface. It has several different terrain types, including craters, plains, and many faults that cross the surface. Miranda has fault canyons and terraced land formations that indicate a diverse tectonic environment. Scientists studying the Voyager 2 images of Triton found evidence of an active geologic past as well as ongoing eruptions of ice volcanoes.

Scientists are still gathering information from the Galileo probe of the Jupiter moon system. Three of Jupiter’s four Galilean satellites show signs of being tectonically active. Europa, Ganymede, and Io all exhibit various features that indicate tectonic motion or volcanism. Europa’s surface is broken apart into large plates similar to the plates found on Earth. The plate movement indicates that the crust is brittle and that the plates move over the top of a softer, more fluid layer. Ganymede probably has a metallic inner core and at least two outer layers that make up a crust and mantle. Io may also have a giant iron core interior that causes the active tectonics and volcanism. It is believed that Io has a partially molten rock mantle and crust. See also Planetary Science: Volcanism and Tectonic Activity.

VI -HISTORY OF TECTONIC THEORY

The theory of plate tectonics arose from several previous geologic theories and discoveries. As early as the 16th century, explorers began examining the coastlines of Africa and South America and proposed that these continents were once connected. In the 20th century, scientists proposed theories that the continents moved or drifted apart from each other. Additionally, in the 1950s scientists proposed that the earth’s magnetic poles wander, leading to more evidence, such as rocks with similar magnetic patterns around the world, that the continents had drifted. More recently, scientists examining the seafloor have discovered that it is spreading as new seafloor is created, and through this work they have discovered that the magnetic polarity of the earth has changed several times throughout the earth's history. The theory of plate tectonics revolutionized earth sciences by providing a framework that could explain these discoveries, as well as events such as earthquakes and volcanic eruptions, mountain building and the formation of the continents and oceans. See also Earthquake.

A -Continental Drift

Beginning in the late 16th century and early 17th century, many people, including Flemish cartographer Abraham Ortelius and English philosopher Sir Francis Bacon, were intrigued by the shapes of the South American and African coastlines and the possibility that these continents were once connected. In 1912, German scientist Alfred Wegener eventually developed the idea that the continents were at one time connected into the theory of continental drift. Scientists of the early 20th century found evidence of continental drift in the similarity of the coastlines and geologic features on both continents. Geologists found rocks of the same age and type on opposite sides of the ocean, fossils of similar animals and plants, and similar ancient climate indicators, such as glaciation patterns. British geologist

Arthur Holmes proposed that convection currents drove the drifting movement of continents. Most earth scientists did not seriously consider the theory of continental drift until the 1960s when scientists began to discover other evidence, such as polar wandering, seafloor spreading, and reversals of the earth’s magnetic field. See also Continent.

B -Polar Wandering

In the 1950s, physicists in England became interested in the observation that certain kinds of rocks produced a magnetic field. They soon decided that the magnetic fields were remnant, or left over, magnetism acquired from the earth's magnetic field as the rocks cooled and solidified from the hot magma that formed them. Scientists measured the orientation and direction of the acquired magnetic fields and, from these orientations, calculated the direction of the rock's magnetism and the distance from the place the rock was found to the magnetic poles. As calculations from rocks of varying ages began to accumulate, scientists calculated the position of the earth's magnetic poles over time. The position of the poles varied depending on where the rocks were collected, and the idea of a polar wander path began to form. When sample paths of polar wander from two continents, such as North America and Europe, were compared, they coincided as if the continents were once joined. This new science and methodology became known as the discipline of paleomagnetism. As a result, discussion of the theory of continental drift increased, but most earth scientists remained skeptical.

C -Seafloor Spreading

During the 1950s, as people began creating detailed maps of the world's ocean floor, they discovered a mid-ocean ridge system of mountains nearly 60,000 km (nearly 40,000 mi) long. This ridge goes all the way around the globe. American geologist Harry H. Hess proposed that this mountain chain was the place where new ocean floor was created and that the continents moved as a result of the expansion of the ocean floors. This process was termed seafloor spreading by American geophysicist Robert S. Dietz in 1961. Hess also proposed that since the size of the earth seems to have remained constant, the seafloor must also be recycled back into the mantle beneath mountain chains and volcanic arcs along the deep trenches on the ocean floor.

These studies also found marine magnetic anomalies, or differences, on the sea floor. The anomalies are changes, or switches, in the north and south polarity of the magnetic rock of the seafloor. Scientists discovered that the switches make a striped pattern of the positive and negative magnetic anomalies: one segment, or stripe, is positive, and the segment next to it is negative. The stripes are parallel to the mid-ocean ridge crest, and the pattern is the same on both sides of that crest. Scientists could not explain the cause of these anomalies until they discovered that the earth's magnetic field periodically reverses direction.

D -Magnetic Field Reversals

In 1963, British scientists Fred J. Vine and Drummond H. Matthews combined their observations of the marine magnetic anomalies with the concept of reversals of the earth's magnetic field. They proposed that the marine magnetic anomalies were a "tape recording" of the spreading of the ocean floor as the earth's magnetic field reversed its direction. At the same time, other geophysicists were studying lava flows from many parts of the world to see how these flows revealed the record of reversals of the direction of the earth's magnetic field. These studies showed that nearly four reversals have occurred over the past 5 million years. The concept of magnetic field reversals was a breakthrough that explained the magnetic polarity switches seen in seafloor spreading as well as the concept of similar magnetic patterns in the rocks used to demonstrate continental drift.

E -Revolution in Geology

The theory of plate tectonics tied together the concepts of continental drift, polar wandering, seafloor spreading, and magnetic field reversals into a single theory that completely changed the science of geology. Geologists finally had one theory that could explain all the different evidence they had accumulated to support these previous theories and discoveries. Geologists now use the theory of plate tectonics to integrate geologic events, to explain the occurrence of earthquakes and volcanic eruptions, and to explain the formation of mountain ranges and oceans.

Nuclear Energy

I –INTRODUCTION

Nuclear Energy, energy released during the splitting or fusing of atomic nuclei. The energy of any system, whether physical, chemical, or nuclear, is manifested by the system's ability to do work or to release heat or radiation. The total energy in a system is always conserved, but it can be transferred to another system or changed in form.

Until about 1800 the principal fuel was wood, its energy derived from solar energy stored in plants during their lifetimes. Since the Industrial Revolution, people have depended on fossil fuels—coal, petroleum, and natural gas—also derived from stored solar energy. When a fossil fuel such as coal is burned, atoms of hydrogen and carbon in the coal combine with oxygen atoms in air. Water and carbon dioxide are produced and heat is released, equivalent to about 1.6 kilowatt-hours per kilogram or about 10 electron volts (eV) per atom of carbon. This amount of energy is typical of chemical reactions resulting from changes in the electronic structure of the atoms. A part of the energy released as heat keeps the adjacent fuel hot enough to keep the reaction going.

II -THE ATOM

The atom consists of a small, massive, positively charged core (nucleus) surrounded by electrons (see Atom). The nucleus, containing most of the mass of the atom, is itself composed of neutrons and protons bound together by very strong nuclear forces, much greater than the electrical forces that bind the electrons to the nucleus. The mass number A of a nucleus is the number of nucleons, or protons and neutrons, it contains; the atomic number Z is the number of positively charged protons. A specific nucleus is designated as ¿U the expression ¯U, for example, represents uranium-235. See Isotope.

The binding energy of a nucleus is a measure of how tightly its protons and neutrons are held together by the nuclear forces. The binding energy per nucleon, the energy required to remove one neutron or proton from a nucleus, is a function of the mass number A. The curve of binding energy implies that if two light nuclei near the left end of the curve coalesce to form a heavier nucleus, or if a heavy nucleus at the far right splits into two lighter ones, more tightly bound nuclei result, and energy will be released. Nuclear energy, measured in millions of electron volts (MeV), is released by the fusion of two light nuclei, as when two heavy hydrogen nuclei, deuterons (ªH), combine in the reaction producing a helium-3 atom, a free neutron (¦n), and 3.2 MeV, or 5.1 × 10-13 J (1.2 × 10-13 cal). Nuclear energy is also released when the fission of a heavy nucleus such as ¯U is induced by the absorption of a neutron as in

producing cesium-140, rubidium-93, three neutrons, and 200 MeV, or 3.2 × 10-11 J (7.7 × 10-12 cal). A nuclear fission reaction releases 10 million times as much energy as is released in a typical chemical reaction. See Nuclear Chemistry.

III -NUCLEAR ENERGY FROM FISSION

The two key characteristics of nuclear fission important for the practical release of nuclear energy are both evident in equation (2). First, the energy per fission is very large. In practical units, the fission of 1 kg (2.2 lb) of uranium-235 releases 18.7 million kilowatt-hours as heat. Second, the fission process initiated by the absorption of one neutron in uranium-235 releases about 2.5 neutrons, on the average, from the split nuclei. The neutrons released in this manner quickly cause the fission of two more atoms, thereby releasing four or more additional neutrons and initiating a self-sustaining series of nuclear fissions, or a chain reaction, which results in continuous release of nuclear energy.

Naturally occurring uranium contains only 0.71 percent uranium-235; the remainder is the nonfissile isotope uranium-238. A mass of natural uranium by itself, no matter how large, cannot sustain a chain reaction because only the uranium-235 is easily fissionable. The probability that a fission neutron with an initial energy of about 1 MeV will induce fission is rather low, but the probability can be increased by a factor of hundreds when the neutron is slowed down through a series of elastic collisions with light nuclei such as hydrogen, deuterium, or carbon. This fact is the basis for the design of practical energy-producing fission reactors.

In December 1942 at the University of Chicago, the Italian physicist Enrico Fermi succeeded in producing the first nuclear chain reaction. This was done with an arrangement of natural uranium lumps distributed within a large stack of pure graphite, a form of carbon. In Fermi's “pile,” or nuclear reactor, the graphite moderator served to slow the neutrons.

IV -NUCLEAR POWER REACTORS

The first large-scale nuclear reactors were built in 1944 at Hanford, Washington, for the production of nuclear weapons material. The fuel was natural uranium metal; the moderator, graphite. Plutonium was produced in these plants by neutron absorption in uranium-238; the power produced was not used.

A -Light-Water and Heavy-Water Reactors

A variety of reactor types, characterized by the type of fuel, moderator, and coolant used, have been built throughout the world for the production of electric power. In the United States, with few exceptions, power reactors use nuclear fuel in the form of uranium oxide isotopically enriched to about three percent uranium-235. The moderator and coolant are highly purified ordinary water. A reactor of this type is called a light-water reactor (LWR).

In the pressurized-water reactor (PWR), a version of the LWR system, the water coolant operates at a pressure of about 150 atmospheres. It is pumped through the reactor core, where it is heated to about 325° C (about 620° F). The superheated water is pumped through a steam generator, where, through heat exchangers, a secondary loop of water is heated and converted to steam. This steam drives one or more turbine generators, is condensed, and is pumped back to the steam generator. The secondary loop is isolated from the water in the reactor core and, therefore, is not radioactive. A third stream of water from a lake, river, or cooling tower is used to condense the steam. The reactor pressure vessel is about 15 m (about 49 ft) high and 5 m (about 16.4 ft) in diameter, with walls 25 cm (about 10 in) thick. The core houses some 82 metric tons of uranium oxide contained in thin corrosion-resistant tubes clustered into fuel bundles.

In the boiling-water reactor (BWR), a second type of LWR, the water coolant is permitted to boil within the core, by operating at somewhat lower pressure. The steam produced in the reactor pressure vessel is piped directly to the turbine generator, is condensed, and is then pumped back to the reactor. Although the steam is radioactive, there is no intermediate heat exchanger between the reactor and turbine to decrease efficiency. As in the PWR, the condenser cooling water has a separate source, such as a lake or river. The power level of an operating reactor is monitored by a variety of thermal, flow, and nuclear instruments. Power output is controlled by inserting or removing from the core a group of neutron-absorbing control rods. The position of these rods determines the power level at which the chain reaction is just self-sustaining.

During operation, and even after shutdown, a large, 1,000-megawatt (MW) power reactor contains billions of curies of radioactivity. Radiation emitted from the reactor during operation and from the fission products after shutdown is absorbed in thick concrete shields around the reactor and primary coolant system. Other safety features include emergency core cooling systems to prevent core overheating in the event of malfunction of the main coolant systems and, in most countries, a large steel and concrete containment building to retain any radioactive elements that might escape in the event of a leak.

Although more than 100 nuclear power plants were operating or being built in the United States at the beginning of the 1980s, in the aftermath of the Three Mile Island accident in Pennsylvania in 1979 safety concerns and economic factors combined to block any additional growth in nuclear power. No orders for nuclear plants have been placed in the United States since 1978, and some plants that have been completed have not been allowed to operate. In 1996 about 22 percent of the electric power generated in the United States came from nuclear power plants. In contrast, in France almost three-quarters of the electricity generated was from nuclear power plants.

In the initial period of nuclear power development in the early 1950s, enriched uranium was available only in the United States and the Union of Soviet Socialist Republics (USSR). The nuclear power programs in Canada, France, and the United Kingdom therefore centered about natural uranium reactors, in which ordinary water cannot be used as the moderator because it absorbs too many neutrons. This limitation led Canadian engineers to develop a reactor cooled and moderated by deuterium oxide (D2O), or heavy water. The Canadian deuterium-uranium reactor known as CANDU has operated satisfactorily in Canada, and similar plants have been built in India, Argentina, and elsewhere.

In the United Kingdom and France the first full-scale power reactors were fueled with natural uranium metal, were graphite-moderated, and were cooled with carbon dioxide gas under pressure. These initial designs have been superseded in the United Kingdom by a system that uses enriched uranium fuel. In France the initial reactor type chosen was dropped in favor of the PWR of U.S. design when enriched uranium became available from French isotope-enrichment plants. Russia and the other successor states of the USSR had a large nuclear power program, using both graphite-moderated and PWR systems.

B -Propulsion Reactors

Nuclear power plants similar to the PWR are used for the propulsion plants of large surface naval vessels such as the aircraft carrier USS Nimitz. The basic technology of the PWR system was first developed in the U.S. naval reactor program directed by Admiral Hyman G. Rickover. Reactors for submarine propulsion are generally physically smaller and use more highly enriched uranium to permit a compact core. The United States, the United Kingdom, Russia, and France all have nuclear-powered submarines with such power plants.

Three experimental seagoing nuclear cargo ships were operated for limited periods by the United States, Germany, and Japan. Although they were technically successful, economic conditions and restrictive port regulations brought an end to these projects. The Soviet government built the first successful nuclear-powered icebreaker, Lenin, for use in clearing the Arctic sea-lanes.

C -Research Reactors

A variety of small nuclear reactors have been built in many countries for use in education and training, research, and the production of radioactive isotopes. These reactors generally operate at power levels near one MW, and they are more easily started up and shut down than larger power reactors.

A widely used type is called the swimming-pool reactor. The core is partially or fully enriched uranium-235 contained in aluminum alloy plates, immersed in a large pool of water that serves as both coolant and moderator. Materials may be placed directly in or near the reactor core to be irradiated with neutrons. Various radioactive isotopes can be produced for use in medicine, research, and industry (see Isotopic Tracer). Neutrons may also be extracted from the reactor core by means of beam tubes to be used for experimentation.

D -Breeder Reactors

Uranium, the natural resource on which nuclear power is based, occurs in scattered deposits throughout the world. Its total supply is not fully known, and may be limited unless sources of very low concentration such as granites and shale were to be used. Conservatively estimated U.S.

resources of uranium having an acceptable cost lie in the range of two million to five million metric tons. The lower amount could support an LWR nuclear power system providing about 30 percent of U.S. electric power for only about 50 years. The principal reason for this relatively brief life span of the LWR nuclear power system is its very low efficiency in the use of uranium: only approximately one percent of the energy content of the uranium is made available in this system.

The key feature of a breeder reactor is that it produces more fuel than it consumes. It does this by promoting the absorption of excess neutrons in a fertile material. Several breeder reactor systems are technically feasible. The breeder system that has received the greatest worldwide attention uses uranium-238 as the fertile material. When uranium-238 absorbs neutrons in the reactor, it is transmuted to a new fissionable material, plutonium, through a nuclear process called β (beta) decay. The sequence of nuclear reactions is In beta decay a nuclear neutron decays into a proton and a beta particle (a high-energy electron).

When plutonium-239 itself absorbs a neutron, fission can occur, and on the average about 2.8 neutrons are released. In an operating reactor, one of these neutrons is needed to cause the next fission and keep the chain reaction going. On the average about 0.5 neutron is uselessly lost by absorption in the reactor structure or coolant. The remaining 1.3 neutrons can be absorbed in uranium-238 to produce more plutonium via the reactions in equation (3).

The breeder system that has had the greatest development effort is called the liquid-metal fast breeder reactor (LMFBR). In order to maximize the production of plutonium-239, the velocity of the neutrons causing fission must remain fast—at or near their initial release energy. Any moderating materials, such as water, that might slow the neutrons must be excluded from the reactor. A molten metal, liquid sodium, is the preferred coolant liquid. Sodium has very good heat transfer properties, melts at about 100° C (about 212° F), and does not boil until about 900° C (about 1650° F). Its main drawbacks are its chemical reactivity with air and water and the high level of radioactivity induced in it in the reactor.

Development of the LMFBR system began in the United States before 1950, with the construction of the first experimental breeder reactor, EBR-1. A larger U.S. program, on the Clinch River, was halted in 1983, and only experimental work was to continue (see Tennessee Valley Authority). In the United Kingdom, France, and Russia and the other successor states of the USSR, working breeder reactors were installed, and experimental work continued in Germany and Japan.

In one design of a large LMFBR power plant, the core of the reactor consists of thousands of thin stainless steel tubes containing mixed uranium and plutonium oxide fuel: about 15 to 20 percent plutonium-239, the remainder uranium. Surrounding the core is a region called the breeder blanket, which contains similar rods filled only with uranium oxide. The entire core and blanket assembly measures about 3 m (about 10 ft) high by about 5 m (about 16.4 ft) in diameter and is supported in a large vessel containing molten sodium that leaves the reactor at about 500° C (about 930° F). This vessel also contains the pumps and heat exchangers that aid in removing heat from the core. Steam is produced in a second sodium loop, separated from the radioactive reactor coolant loop by the intermediate heat exchangers in the reactor vessel. The entire nuclear reactor system is housed in a large steel and concrete containment building.

The first large-scale plant of this type for the generation of electricity, called Super-Phénix, went into operation in France in 1984. (However, concerns about operational safety and environmental contamination led the French government to announce in 1998 that Super-Phénix would be dismantled). An intermediate-scale plant, the BN-600, was built on the shore of the Caspian Sea for the production of power and the desalination of water. The British have a large 250-MW prototype in Scotland.

The LMFBR produces about 20 percent more fuel than it consumes. In a large power reactor enough excess new fuel is produced over 20 years to permit the loading of another similar reactor. In the LMFBR system about 75 percent of the energy content of natural uranium is made available, in contrast to the one percent in the LWR.

V -NUCLEAR FUELS AND WASTES

The hazardous fuels used in nuclear reactors present handling problems in their use. This is particularly true of the spent fuels, which must be stored or disposed of in some way.

A -The Nuclear Fuel Cycle

Any electric power generating plant is only one part of a total energy cycle. The uranium fuel cycle that is employed for LWR systems currently dominates worldwide nuclear power production and includes many steps. Uranium, which contains about 0.7 percent uranium-235, is obtained from either surface or underground mines. The ore is concentrated by milling and then shipped to a conversion plant, where its elemental form is changed to uranium hexafluoride gas (UF6). At an isotope enrichment plant, the gas is forced against a porous barrier that permits the lighter uranium-235 to penetrate more readily than uranium-238. This process enriches uranium to about 3 percent uranium-235. The depleted uranium—the tailings—contain about 0.3 percent uranium-235.

The enriched product is sent to a fuel fabrication plant, where the UF6 gas is converted to uranium oxide powder, then into ceramic pellets that are loaded into corrosion-resistant fuel rods. These are assembled into fuel elements and are shipped to the reactor power plant. The world's supply of enriched uranium fuel for powering commercial nuclear power plants is produced by five consortiums located in the United States, Western Europe, Russia, and Japan. The United States consortium—the federally owned United States Enrichment Corporation—produces 40 percent of this enriched uranium.

A typical 1,000-MW pressurized-water reactor has about 200 fuel elements, one-third of which are replaced each year because of the depletion of the uranium-235 and the buildup of fission products that absorb neutrons. At the end of its life in the reactor, the fuel is tremendously radioactive because of the fission products it contains and hence is still producing a considerable amount of energy. The discharged fuel is placed in water storage pools at the reactor site for a year or more.

At the end of the cooling period the spent fuel elements are shipped in heavily shielded casks either to permanent storage facilities or to a chemical reprocessing plant. At a reprocessing plant, the unused uranium and the plutonium-239 produced in the reactor are recovered and the radioactive wastes concentrated. (In the late 1990s neither such facility was yet available in the United States for power plant fuel, and temporary storage was used.)

The spent fuel still contains almost all the original uranium-238, about one-third of the uranium-235, and some of the plutonium-239 produced in the reactor. In cases where the spent fuel is sent to permanent storage, none of this potential energy content is used. In cases where the fuel is reprocessed, the uranium is recycled through the diffusion plant, and the recovered plutonium-239 may be used in place of some uranium-235 in new fuel elements. At the end of the 20th century, no reprocessing of fuel occurred in the United States because of environmental, health, and safety concerns, and the concern that plutonium-239 could be used illegally for the manufacture of weapons.

In the fuel cycle for the LMFBR, plutonium bred in the reactor is always recycled for use in new fuel. The feed to the fuel-element fabrication plant consists of recycled uranium-238, depleted uranium from the isotope separation plant stockpile, and part of the recovered plutonium-239. No additional uranium needs to be mined, as the existing stockpile could support many breeder reactors for centuries. Because the breeder produces more plutonium-239 than it requires for its own refueling, about 20 percent of the recovered plutonium is stored for later use in starting up new breeders. Because new fuel is bred from the uranium-238, instead of using only the natural uranium-235 content, about 75 percent of the potential energy of uranium is made available with the breeder cycle.

The final step in any of the fuel cycles is the long-term storage of the highly radioactive wastes, which remain biologically hazardous for thousands of years. Fuel elements may be stored in shielded, guarded repositories for later disposition or may be converted to very stable compounds, fixed in ceramics or glass, encapsulated in stainless steel canisters, and buried far underground in very stable geologic formations.

However, the safety of such repositories is the subject of public controversy, especially in the geographic region in which the repository is located or is proposed to be built. For example, environmentalists plan to file a lawsuit to close a repository built near Carlsbad, New Mexico. In 1999, this repository began receiving shipments of radioactive waste from the manufacture of nuclear weapons in United States during the Cold War. Another controversy centers around a proposed repository at Yucca Mountain, Nevada. Opposition from state residents and questions about the geologic stability of this site have helped prolong government studies. Even if opened, the site will not receive shipments of radioactive waste until at least 2010 (see Nuclear Fuels and Wastes, Waste Management section below).

B -Nuclear Safety

Public concern about the acceptability of nuclear power from fission arises from two basic features of the system. The first is the high level of radioactivity present at various stages of the nuclear cycle, including disposal. The second is the fact that the nuclear fuels uranium-235 and plutonium-239 are the materials from which nuclear weapons are made. See Nuclear Weapons; Radioactive Fallout.

U.S. President Dwight D. Eisenhower announced the U.S. Atoms for Peace program in 1953. It was perceived as offering a future of cheap, plentiful energy. The utility industry hoped that nuclear power would replace increasingly scarce fossil fuels and lower the cost of electricity. Groups concerned with conserving natural resources foresaw a reduction in air pollution and strip mining. The public in general looked favorably on this new energy source, seeing the program as a realization of hopes for the transition of nuclear power from wartime to peaceful uses.

Nevertheless, after this initial euphoria, reservations about nuclear energy grew as greater scrutiny was given to issues of nuclear safety and weapons proliferation. In the United States and other countries many groups oppose nuclear power. In addition, high construction costs, strict building and operating regulations, and high costs for waste disposal make nuclear power plants much more expensive to build and operate than plants that burn fossil fuels. In some industrialized countries, the nuclear power industry has come under growing pressure to cut operating expenses and become more cost-competitive. Other countries have begun or planned to phase out nuclear power completely.

At the end of the 20th century, many experts viewed Asia as the only possible growth area for nuclear power. In the late 1990s, China, Japan, South Korea, and Taiwan had nuclear power plants under construction. However, many European nations were reducing or reversing their commitments to nuclear power. For example, Sweden committed to phasing out nuclear power by 2010. France canceled several planned reactors and was considering the replacement of aging nuclear plants with environmentally safer fossil-fuel plants. Germany announced plans in 1998 to phase out nuclear energy. In the United States, no new reactors had been ordered since 1978.

In 1996, 21.9 percent of the electricity generated in the United States was produced by nuclear power. By 1998 that amount had decreased to 20 percent. Because no orders for nuclear plants have been placed since 1978, this share should continue to decline as existing nuclear plants are eventually closed. In 1998 Commonwealth Edison, the largest private owner and operator of nuclear plants in the United States, had only four of 12 nuclear power plants online. Industry experts cite economic, safety, and labor problems as reasons for these shutdowns.

B1 -Radiological Hazards

Radioactive materials emit penetrating, ionizing radiation that can injure living tissues. The commonly used unit of radiation dose equivalent in humans is the sievert. (In the United States, rems are still used as a measure of dose equivalent. One rem equals 0.01 sievert.) Each individual in the United States and Canada is exposed to about 0.003 sievert per year from natural background radiation sources. An exposure to an individual of five sieverts is likely to be fatal. A large population exposed to low levels of radiation will experience about one additional cancer for each 10 sieverts total dose equivalent. See Radiation Effects, Biological.

Radiological hazards can arise in most steps of the nuclear fuel cycle. Radioactive radon gas is a colorless gas produced from the decay of uranium. As a result, radon is a common air pollutant in underground uranium mines. The mining and ore-milling operations leave large amounts of waste material on the ground that still contain small concentrations of uranium. To prevent the release of radioactive radon gas into the air from this uranium waste, these wastes must be stored in waterproof basins and covered with a thick layer of soil.

Uranium enrichment and fuel fabrication plants contain large quantities of three-percent uranium-235, in the form of corrosive gas, uranium hexafluoride, UF6. The radiological hazard, however, is low, and the usual care taken with a valuable material posing a typical chemical hazard suffices to ensure safety.

B2 -Reactor Safety Systems

The safety of the power reactor itself has received the greatest attention. In an operating reactor, the fuel elements contain by far the largest fraction of the total radioactive inventory. A number of barriers prevent fission products from leaking into the air during normal operation. The fuel is clad in corrosion-resistant tubing. The heavy steel walls of the primary coolant system of the PWR form a second barrier. The water coolant itself absorbs some of the biologically important radioactive isotopes such as iodine. The steel and concrete building is a third barrier.

During the operation of a power reactor, some radioactive compounds are unavoidably released. The total exposure to people living nearby is usually only a few percent of the natural background radiation. Major concerns arise, however, from radioactive releases caused by accidents in which fuel damage occurs and safety devices fail. The major danger to the integrity of the fuel is a loss-of-coolant accident in which the fuel is damaged or even melts. Fission products are released into the coolant, and if the coolant system is breached, fission products enter the reactor building.

Reactor systems rely on elaborate instrumentation to monitor their condition and to control the safety systems used to shut down the reactor under abnormal circumstances. Backup safety systems that inject boron into the coolant to absorb neutrons and stop the chain reaction to further assure shutdown are part of the PWR design. Light-water reactor plants operate at high coolant pressure. In the event of a large pipe break, much of the coolant would flash into steam and core cooling could be lost. To prevent a total loss of core cooling, reactors are provided with emergency core cooling systems that begin to operate automatically on the loss of primary coolant pressure. In the event of a steam leak into the containment building from a broken primary coolant line, spray coolers are actuated to condense the steam and prevent a hazardous pressure rise in the building.

B3 -Three Mile Island and Chernobyl'

Despite the many safety features described above, an accident did occur in 1979 at the Three Mile Island PWR near Harrisburg, Pennsylvania. A maintenance error and a defective valve led to a loss-of-coolant accident. The reactor itself was shut down by its safety system when the accident began, and the emergency core cooling system began operating as required a short time into the accident. Then, however, as a result of human error, the emergency cooling system was shut off, causing severe core damage and the release of volatile fission products from the reactor vessel. Although only a small amount of radioactive gas escaped from the containment building, causing a slight rise in individual human exposure levels, the financial damage to the utility was very large, $1 billion or more, and the psychological stress on the public, especially those people who live in the area near the nuclear power plant, was in some instances severe.

The official investigation of the accident named operational error and inadequate control room design, rather than simple equipment failure, as the principal causes of the accident. It led to enactment of legislation requiring the Nuclear Regulatory Commission to adopt far more stringent standards for the design and construction of nuclear power plants. The legislation also required utility companies to assume responsibility for helping state and county governments prepare emergency response plans to protect the public health in the event of other such accidents.

Since 1981, the financial burdens imposed by these requirements have made it difficult to build and operate new nuclear power plants. Combined with other factors, such as high capital costs and long construction periods (which means builders must borrow more money and wait longer periods before earning a return on their investment), safety regulations have forced utility companies in the states of Washington, Ohio, Indiana, and New York to abandon partly completed plants after spending billions of dollars on them.
On April 26, 1986, another serious incident alarmed the world. One of four nuclear reactors at Chernobyl', near Pripyat', about 130 km (about 80 mi) north of Kyiv (now in Ukraine) in the USSR, exploded and burned. Radioactive material spread over Scandinavia and northern Europe, as discovered by Swedish observers on April 28. According to the official report issued in August, the accident was caused by unauthorized testing of the reactor by its operators. The reactor went out of control; there were two explosions, the top of the reactor blew off, and the core was ignited, burning at temperatures of 1500° C (2800° F). Radiation about 50 times higher than that at Three Mile Island exposed people nearest the reactor, and a cloud of radioactive fallout spread westward. Unlike most reactors in western countries, including the United States, the reactor at Chernobyl' did not have a containment building. Such a structure could have prevented material from leaving the reactor site. About 135,000 people were evacuated, and more than 30 died. The plant was encased in concrete. By 1988, however, the other three Chernobyl' reactors were back in operation. One of the three remaining reactors was shut down in 1991 because of a fire in the reactor building. In 1994 Western nations developed a financial aid package to help close the entire plant, and a year later the Ukrainian government finally agreed to a plan that would shut down the remaining reactors by the year 2000.

C -Fuel Reprocessing

The fuel reprocessing step poses a combination of radiological hazards. One is the accidental release of fission products if a leak should occur in chemical equipment or the cells and building housing it. Another may be the routine release of low levels of inert radioactive gases such as xenon and krypton. In 1966 a commercial reprocessing plant opened in West Valley, New York. But in 1972 this reprocessing plant was closed after generating more than 600,000 gallons of high-level radioactive waste. After the plant was closed, a portion of this radioactive waste was partially treated and cemented into nearly 20,000 steel drums. In 1996, the United States Department of Energy began to solidify the remaining liquid radioactive wastes into glass cylinders. At the end of the 20th century, no reprocessing plants were licensed in the United States.

Of major concern in chemical reprocessing is the separation of plutonium-239, a material that can be used to make nuclear weapons. The hazards of theft of plutonium-239, or its use for intentional but hidden production for weapons purposes, can best be controlled by political rather than technical means. Improved security measures at sensitive points in the fuel cycle and expanded international inspection by the International Atomic Energy Agency (IAEA) offer the best prospects for controlling the hazards of plutonium diversion.

D -Waste Management

The last step in the nuclear fuel cycle, waste management, remains one of the most controversial. The principal issue here is not so much the present danger as the danger to generations far in the future. Many nuclear wastes remain radioactive for thousands of years, beyond the span of any human institution. The technology for packaging the wastes so that they pose no current hazard is relatively straightforward. The difficulty lies both in being adequately confident that future generations are well protected and in making the political decision on how and where to proceed with waste storage. Permanent but potentially retrievable storage in deep stable geologic formations seems the best solution. In 1988 the U.S. government chose Yucca Mountain, a Nevada desert site with a thick section of porous volcanic rocks, as the nation's first permanent underground repository for more than 36,290 metric tons of nuclear waste. However, opposition from state residents and uncertainty that Yucca Mountain may not be completely insulated from earthquakes and other hazards has prolonged government studies. For example, a geological study by the U.S. Department of Energy detected water in several mineral samples taken at the Yucca Mountain site. The presence of water in these samples suggests that water may have once risen up through the mountain and later subsided. Because such an event could jeopardize the safety of a nuclear waste repository, the Department of Energy has funded more study of these fluid intrusions.

A $2 billion repository built in underground salt caverns near Carlsbad, New Mexico, is designed to store radioactive waste from the manufacture of nuclear weapons during the Cold War. This repository, located 655 meters (2,150 feet) underground, is designed to slowly collapse and encapsulate the plutonium-contaminated waste in the salt beds. Although the repository began receiving radioactive waste shipments in April 1999, environmentalists planned to file a lawsuit to close the Carlsbad repository.

VI -NUCLEAR FUSION

The release of nuclear energy can occur at the low end of the binding energy curve (see accompanying chart) through the fusion of two light nuclei into a heavier one. The energy radiated by stars, including the Sun, arises from such fusion reactions deep in their interiors. At the enormous pressure and at temperatures above 15 million ° C (27 million ° F) existing there, hydrogen nuclei combine according to equation (1) and give rise to most of the energy released by the Sun.

Nuclear fusion was first achieved on earth in the early 1930s by bombarding a target containing deuterium, the mass-2 isotope of hydrogen, with high-energy deuterons in a cyclotron (see Particle Accelerators). To accelerate the deuteron beam a great deal of energy is required, most of which appeared as heat in the target. As a result, no net useful energy was produced. In the 1950s the first large-scale but uncontrolled release of fusion energy was demonstrated in the tests of thermonuclear weapons by the United States, the USSR, the United Kingdom, and France. This was such a brief and uncontrolled release that it could not be used for the production of electric power.

In the fission reactions discussed earlier, the neutron, which has no electric charge, can easily approach and react with a fissionable nucleus—for example, uranium-235. In the typical fusion reaction, however, the reacting nuclei both have a positive electric charge, and the natural repulsion between them, called Coulomb repulsion, must be overcome before they can join. This occurs when the temperature of the reacting gas is sufficiently high—50 to 100 million ° C (90 to 180 million ° F). In a gas of the heavy hydrogen isotopes deuterium and tritium at such temperature, the fusion reaction occurs, releasing about 17.6 MeV per fusion event. The energy appears first as kinetic energy of the helium-4 nucleus and the neutron, but is soon transformed into heat in the gas and surrounding materials.

If the density of the gas is sufficient—and at these temperatures the density need be only 10-5 atm, or almost a vacuum—the energetic helium-4 nucleus can transfer its energy to the surrounding hydrogen gas, thereby maintaining the high temperature and allowing subsequent fusion reactions, or a fusion chain reaction, to take place. Under these conditions, “nuclear ignition” is said to have occurred.

The basic problems in attaining useful nuclear fusion conditions are (1) to heat the gas to these very high temperatures and (2) to confine a sufficient quantity of the reacting nuclei for a long enough time to permit the release of more energy than is needed to heat and confine the gas. A subsequent major problem is the capture of this energy and its conversion to electricity.

At temperatures of even 100,000° C (180,000° F), all the hydrogen atoms are fully ionized. The gas consists of an electrically neutral assemblage of positively charged nuclei and negatively charged free electrons. This state of matter is called a plasma.

A plasma hot enough for fusion cannot be contained by ordinary materials. The plasma would cool very rapidly, and the vessel walls would be destroyed by the extreme heat. However, since the plasma consists of charged nuclei and electrons, which move in tight spirals around the lines of force of strong magnetic fields, the plasma can be contained in a properly shaped magnetic field region without reacting with material walls.

In any useful fusion device, the energy output must exceed the energy required to confine and heat the plasma. This condition can be met when the product of confinement time t and plasma density n exceeds about 1014. The relationship tn≥ 1014 is called the Lawson criterion.

Numerous schemes for the magnetic confinement of plasma have been tried since 1950 in the United States, Russia, the United Kingdom, Japan, and elsewhere. Thermonuclear reactions have been observed, but the Lawson number rarely exceeded 1012. One device, however—the tokamak, originally suggested in the USSR by Igor Tamm and Andrey Sakharov—began to give encouraging results in the early 1960s.

The confinement chamber of a tokamak has the shape of a torus, with a minor diameter of about 1 m (about 3.3 ft) and a major diameter of about 3 m (about 9.8 ft). A toroidal (donut-shaped) magnetic field of about 50,000 gauss is established inside this chamber by large electromagnets. A longitudinal current of several million amperes is induced in the plasma by the transformer coils that link the torus. The resulting magnetic field lines, spirals in the torus, stably confine the plasma.

Based on the successful operation of small tokamaks at several laboratories, two large devices were built in the early 1980s, one at Princeton University in the United States and one in the USSR. The enormous magnetic fields in a tokamak subject the plasma to extremely high temperatures and pressures, forcing the atomic nuclei to fuse. As the atomic nuclei are fused together, an extraordinary amount of energy is released. During this fusion process, the temperature in the tokamak reaches three times that of the Sun's core.

Another possible route to fusion energy is that of inertial confinement. In this concept, the fuel—tritium or deuterium—is contained within a tiny glass sphere that is then bombarded on several sides by a pulsed laser or heavy ion beam. This causes an implosion of the glass sphere, setting off a thermonuclear reaction that ignites the fuel. Several laboratories in the United States and elsewhere are currently pursuing this possibility. In the late 1990s, many researchers concentrated on the use of beams of heavy ions, such as barium ions, rather than lasers to trigger inertial-confinement fusion. Researchers chose heavy ion beams because heavy ion accelerators can produce intense ion pulses at high repetition rates and because heavy ion accelerators are extremely efficient at converting electric power into ion beam energy, thus reducing the amount of input power. Also in comparison to laser beams, ion beams can penetrate the glass sphere and fuel more effectively to heat the fuel.

Progress in fusion research has been promising, but the development of practical systems for creating a stable fusion reaction that produces more power than it consumes will probably take decades to realize. The research is expensive, as well. However, some progress was made in the early 1990s. In 1991, for the first time ever, a significant amount of energy—about 1.7 million watts—was produced from controlled nuclear fusion at the Joint European Torus (JET) Laboratory in England. In December 1993, researchers at Princeton University used the Tokamak Fusion Test Reactor to produce a controlled fusion reaction that output 5.6 million watts of power. However, both the JET and the Tokamak Fusion Test Reactor consumed more energy than they produced during their operation.

If fusion energy does become practical, it offers the following advantages: (1) a limitless source of fuel, deuterium from the ocean; (2) no possibility of a reactor accident, as the amount of fuel in the system is very small; and (3) waste products much less radioactive and simpler to handle than those from fission systems.

Q1. WHAT QUANTITIES ARE MEASURED BY THE FOLLOWING UNITS?

WAT……………………………………………………….
COLOUMB……………………………………………..
PASCAL…………………………………………………
OHM……………………………………………………..
KELVIN…………………………………………………
JOULE……………………………………………………..
METER……………………………………………………
FARADAY……………………………………………….
HERTZ……………………………………………………
AMPERE…………………………………………………
TORR………………………………………………………
CURIE………………………………………………………
ANGSTROM……………………………………………….
LIGHT…………………………………………………
DIOPTR…………………………………………………….
HORSE…………………………………………….
RADIAN……………………………………………………
CANDELA………………………………………………….
MOLE……………………………………………………….
WEBER……………………………………………………
TESLA……………………………………………………
SIEMEN……………………………………………………
RUTHERFORD…………………………………………
PARSEC……………………………………………………
DEGREE……………………………………………………
STERADIAN………………………………………………
BARREL……………………………………………………
BTU…………………………………………………………
KWH……………………………………………………….
NEWTON…………………………………………………
BECQUEREL………………………………………………
VOLT……………………………………………………….
ACRE-FOOT………………………………………………
CUSEC……………………………………………………..
HERTZ……………………………………………………….
MhO………………………………………………………….
WALT………………………………………………………
DYNE……………………………………………………............
TON..

Q2. WHAT DO THE FOLLOWING ABBREVIATIONS STAND FOR?

LASER……………………………………………………
RADAR………………………………………………
LPG……………………………………………………
CFC………………………………………………………
AIDS……………………………………………………
ROM………………………………………………………
LAN……………………………………………………
WWW……………………………………………………¬
DNA……………………………………………………
HDL………………………………………………………¬
MeV………………………………………………………
UHF………………………………………………………¬
LED……………………………………………………
LCD………………………………………………………¬
BASIC……………………………………………………
MASER……………………………………………………¬
SONAR…………………………………………………
SARS……………………………………………………¬
NTP……………………………………………………
RQ………………………………………………………¬
PVC……………………………………………………
NPN………………………………………………………¬
WAN……………………………………………………
ECG………………………………………………………¬
BCG……………………………………………………
ETT………………………………………………………¬
DBS……………………………………………………
BTU………………………………………………………¬
TNT……………………………………………………
CNG………………………………………………………¬