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Editorial

India ranks 21st in terms of output of research papers in science, but 119th in terms of research papers of any worthwhile contribution. The number of R&D scientists and engineers per million of the population is 157 in India. This is one fifth the ratio in South Korea, and one thirtieth the ratio of countries like the USA and Japan. We have 17 per cent of the global population but account for a mere 1.5 per cent of the global output in R&D. There is no shortage of brilliant scientific minds in India, dedication and application. Our inability to harness our huge pool of scientific talent to rise above mediocrity is due largely to our failure to build up institutions which can nurture and inspire genius. Apart from the IITs, no Indian educational institution merits a place in the list of internationally recognised universities and research centres.

One reason for this is, of course, funds. We spend a mere 3.5 per cent of our GDP on education compared to double this percentage in most countries. Which explains why our HRD Minister Kapil Sibal is in no position to fulfil the just demands of IIT professors that their emoluments should be commensurate to their services and their standing in the international market.

But, probably more than lack of funds, the responsibility for the sorry state of our research institutions is the work culture we foster. Political interference, nepotism, petty jealousies, corruption and bureaucratization have all taken a toll on our institutions of higher education. Merit is seldom the criteria for advancement, particularly as most of our research centres are controlled by government, where red tapism is rewarded and individual initiative frowned upon. Take for instance, the case of the Nobel Prize winner's father, C V Ramakrishnan. He too was a brilliant academic who, together with his wife, founded the Department of Biochemistry at MSU in Vadodara.

Nevertheless, Ramakrishnan was hounded out of the institution because some well connected Ph.D students backed by politicians complained against him. Ramakrishnan's fault was that he went strictly by merit in awarding scholarships, and tried to give research grants to as many students as possible, even if it meant cutting down the amount for each one. Our research institutions are replete with similar examples. Over the years, bureaucratic and political interference has driven out a sizable section of the country's top medical talent from the prestigious All India Institute of Medical Sciences.

—Editors

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Searching for Treatment of Veterinary Diseases from Plants Growing Walls

Anurag Kumar & H. K. Verma

(PG Dept. of Botany JPU, Chapra)

Introduction

Some plants, having certain chemicals are used for the treatment of eases of cattles. During present work an attempt has been made to sort out the plants growing on walls which are useful in the treatment of various.

Veterinary science was developed in India as early as the vedic period Atharvayeda (3500-500BC) has reference to dairy farming, cattle health care etc. The fact that the veterinary practice was flourishing in the vedic age (2000BC) is evidenced in the hymns of 'Atharvaveda' and 'Rigveda'. Nakula and Sahadeva of the epic period were famous, healers of ailments of cattles.1

(i) Nakula Samhita - Treatise of treatment of cattle by Nakula and Sahadeva.

(ii) Answayurveda-Treatise on treatment of horses by Shalihotra (Father of Vetermary Science.)

(iii) Hasthyurveda- Treatise on treatment of elephants by Palakopya.

Megasthanes the Greek Ambatssador in India during 400BC was highly impressed with what he saw of veterinary practices, Ashoka, the Buddhist emperor (300BC) established a network of veterinary hospitals throughout India. References on animal husbandry are available in Kautilya's Arthrastra' (200BC). Similar technology was brought along by Muslim invaders who settled in India. Both Ayurvedic and Unani systems were in practice in India until 1800, at the time of British take over.

There has been a rich tradition and indigenous knowledge about animal Health care in India. However, during the last hundred years or so, there has been a decline in the practice of Ayurveda and Unani system of veterinary medicine in preference to wester n veterinary practice based on chemical and synthetic drugs. In rural areas remedies bisect on locally available herbs and animals products are still prevalent.

Recently there has been a revival of these two systems based on herbat medicines. During present work an attempt has been made to sort out the list of plants used in the treatment of different diseases of animals.

Materials and Methods

Several visits of different places have been made to collect the plants growing on walls having medicinal values for animals. The literatures available on medicinal plants were gathered.

The Studies on Flora of Saran (Verma, H.K. 1983) was consulted for identification of the plants and some rural vaidyas were consulted regarding uses of different parts of plants; used.

Observation

A check list of medicinal plants have been prepared. The botanical names of plants growing on walls along with families having medicinal value for animals are listed below with the parts used for the .specific diseases.

1. *Calculus hirsutus (Linn) Diels*: Family Menispermaceae- Leaves are used to remove lice, repel exparasite, In diarrhoea, urinary disease and galactoma. Roots are used in mouth disease and shoot is used in galactagogue.
2. *Tinospora Cordifolia (Willd):* Mierb- Family Menispermaceae - Whole plant is used in indigestion. Roots are used in immunisation and ulcer. Rhizome is used in abdominal pain, fever and as tonic. Stem is used in haematuria, abdominal pain diuretic, vermifuge blood purifier, cardiac stimulant, cough, appetiser and in Jaundice. Bark is used as an thelmintic, diuretic, vermifugal, blood purifier and cardiac stimulant.
3. *Argemone mexicana (Linn.):* Family Pepaveraceae- Whole plant is used as antiseptic and seeds are used to care fungal infection.
4. *Cleome viscosa Linn*: Family Cleomaceae. Whole plant is used in wound.
5. *Cleome gynandra Linn*: Family Cleomacae-Aerial parts are used in anthelmintic vermifugal and Rheumatism and leaves are used in wound and ear.
6. *Abutilon indicum (Linn.) S.W.*: Family Malvaceae- Leaves are used in releif from lice and in eye trouble.
7. *Sida cordifolia Linn.*: Family Malvaceae - Whole plant is used in shivering diseuuse.
8. *Bomhax ceiba Lirrt.:* Family Bombacaceae - Stem bark is used in dislocated bones, vaginal diseases, delivery and in diarrhoea.
9. *Oxalis corniculata Linn.*: Family Oxalidaceae - Whole plants used in skin disuse, scabie, wart and to rinderpest.
10. *Azadirachta indica A. Juss. Family Meliaceae*: Leaves are used in constipation days pepsia, repel exparasite, ulcer, wound skin, stomatitis, prolapse uterus throat cough, indigestion. Fruits are used in cough,; cut and in anthelmintic. Oil is used in toot and Mouth disease and stem bark is used in fever.
11. *Cocclnia grandis (Linn.) Voigt*: Family Cucurbitaceae- Leaves are used in simla, hy groma, swell, galactagougue, cough, cold and fever. Fruits are used in hygroma, swell and dizziness.
12. *Tridax procumbens Linn:* Family Asteraceae - Leaves are used in cut, wound and haemorrhage.
13. *Pulicaria crispa (Forsk.) Olive:* Family Asteraceae- Whole plant is used in brurse and wound.
14. *Eclipta alba (Linn.) Harsk:* Family Asteraceae- Whole plant is used in fever wound *swell neck. Roots in ulcer, antiseptic and in wound. Leaves are used in wound and in* ear trouble,-Latex is used in wound.
15. *Calotropls procera (Wilid.) Dry ex. Alt*: Family Asolepladaceuc-Roots 4.110 used III pyrexia neck injurles. Latex is used in cut. Leaves are used in skin, wound, bronchitis and in mouth.
16. *Lantana camana Linn. Var. aculeata (Linn.) Moldenke Family Verbenceae:* Leaves are used in dyspepsia and in Jaundice.

17. *Oclmum sanctum Linn:* Family Lamiaceae Leaves are used in dyspepsia and seeds are used in boils.
18. *Boerhavia diffusa Linn: Family Nyctaginaceae*: Whole plants is used in diuretic. Roots in black quarter, emetic and expectorant. Leaves are used in eye trouble and aerial parts as stimulant.
19. *Alternanthera sesilis (Linn.) D.G.:* Family Amaranthaceae - Roots are used in ulcer, pneumonia and in peptic ulcer. Whole plant is used in galactagogue.
20. *Celosla cristata Linn:* Family Amaranthaceae - Leaves are used in wound and in burn.
21. *Achyranthus aspera Linn*: Family Amananthaceae- Whole plant is used in diuretic, *eye,* removal of placenta. Aerial-part is used in diuretic. Roots are used iii delivery, bronchitis, gastric, appetiser and cramps. Leaves are used in fever and in injunes.
22. *Amaranthus viridis Linn:* Family Amaranthaceae- Whole plant is, used in galacta gogue.
23. *Chenopodium album Linn*: Family Chenopodiaceae - Leaves are used in cut, wound, *sore, and galactagogue. grains. are used in delivery.*
24. *Basella alba Linn:* Family Basellaceae- Stem is used in wound and in labour pain.
25. *Euphorbia hirta Linn:* Family Euphorbiaceae - Whole plant is used in bone fracture Leaves are used in cholera.
26. *Acalypha indica Linn*: Family Euphorbiaceae - Leaves are used in eye disease and in scabies.
27. *Cannabis sativa Linn:* Family Cannabinaceae - Leaves are used in diarrhoea an the imintic, flatulence and indigestion. Seeds are used in blood dysentery.
28. *Ficus benghalensis Linn:* Family Moraceae- Roots are used in bone tracture and in gastric. Latex is used in neck, injuries.
29. *Ficus religiosa Linn:* Family Moraceae - Leaves are used in broncnuis . Bark is used in wound, in wound, in constipation and in cough.
30. *Ficus racemosa Linn:* Family Moraceae - Leaves are used in is used in render pest and in plague.
31. *Commalina benghalensis Linn:* Family Commelinaceae - Leaves are used in yoke sore. Seeds are used in galactagogue.
32. *Cyperus rotundus Linn:* Family Cyperaceae - Tubers are used in gargati and root is used in expel worms.
33. *Cynodon ductylon (Linn.) Pors:* Family Pancono-Shoots are used in dyspopara.

Rosult and Discussion

The present study reveals that the flora growing on walls of Chapra town have a great potential of ethno-medicinal value. The present investigation has been carried out specifically targeting the phar-maceutical importance of the wall flora for animals ailments.

It has been observed that there are many plants growing widely on .walls of Chapra town having medicinal value for animals as well as humans and are used by the indigenous people since long with confidence as it is acceptable, effective cheap easily available and accessible with no side effect.

Their pharrriaceuhcal knoweledge' still confined to few persons who are specialized in indigenous herbal medicines. It needs proper campaigning and awareness among general people for their agro-tech niques so that these plants could be available in the large number and be protected from becoming endangered species.

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The Organic Constituents of Plants

Dr. Jai Kumar Shah

Assistant

When the water naturally existing in plants is expelled by exposure to the air or a gentle heat, the residual dry matter is found to be composed of a considerable number of different substances, which have been divided into two great classes, called the organic and the inorganic, or mineral constituents of plants. The former are readily combustible, and on the application of heat, catch fire, and are entirely consumed, leaving the inorganic matters in the form of a white residuum or ash. All plants contain both classes of substances; and though their relative proportions vary within very wide limits, the former always greatly exceed the latter, which in many cases form only a very minute proportion of the whole weight of the plant. Owing to the great preponderance of the organic or combustible matters, it was at one time believed that the inorganic substances formed no part of the true structure of plants, and consisted only of a small portion of the mineral matters of the soil, which had been absorbed along with their organic food; but this opinion, which probably was never universally entertained, is now entirely abandoned, and it is no longer doubted that both classes of substances are equally essential to their existence.

Although they form so large a proportion of the plant, its organic constituents are composed of no more than four elements, viz.:

Carbon.

Hydrogen.

Nitrogen.

Oxygen.

The inorganic constituents are much more numerous, not less than thirteen substances, which appear to be essential, having been observed. These are:

Potash.	Soda.
Lime.	Magnesia.
Peroxide of Iron.	Silicic Acid.
Phosphoric Acid.	Sulphuric Acid.
Chlorine.	And more rarely
Manganese.	Iodine.
Bromine.	Fluorine.

Several other substances, among which may be mentioned alumina and copper, have also been enumerated; but there is every reason to believe that they are not essential, and the cases in which they have been found are quite exceptional.

It is to be especially noticed that none of these substances occur in plants in the free or uncombined state, but always in the form of compounds of greater or less complexity, and extremely varied both in their properties and composition.

It would be out of place, in a work like the present, to enter into complete details of the properties of the elements of which plants are composed, which belongs strictly to pure chemistry, but it is necessary to premise a few observations regarding the organic elements, and their more important compounds.

Carbon.—When a piece of wood is heated in a close vessel, it is charred, and converted into charcoal. This charcoal is the most familiar form of carbon, but it is not absolutely pure, as it necessarily contains the ash of the wood from which it was made. In its purest form it occurs in the diamond, which is believed to be produced by the decomposition of vegetable matters, and it is there crystallized and remarkably transparent; but when produced by artificial processes, carbon is always black, more or less porous, and soils the fingers. It is insoluble in water, burns readily, and is converted into carbonic acid. Carbon is the largest constituent of plants, and forms, in round numbers, about 50 per cent of their weight when dry.

Carbonic Acid.—This, the most important compound of carbon and oxygen, is best obtained by pouring a strong acid upon chalk or limestone, when it escapes with effervescence. It is a colourless gas, extinguishing flame, incapable of supporting respiration, much heavier than atmospheric air, and slightly soluble in water, which takes up its own volume of the gas. It is produced abundantly when vegetable matters are burnt, as also during respiration, fermentation, and many other processes. It is likewise formed daring the decay of animal and vegetable matters, and is consequently evolved from dung and compost heaps.

Hydrogen occurs in nature only in combination. Its principal compound is water, from which it is separated by the simultaneous action of an acid, such as sulphuric acid and a metal, in the form of a transparent gas, lighter than any other substance. It is very combustible, burns with a pale blue flame, and is converted into water. It is found in all plants, although in comparatively small quantity, for, when dry, they rarely contain more than four or five per cent. Its most important compound is water, of which it forms one-ninth, the other eight-ninths consisting of oxygen.

Nitrogen exists abundantly in the atmosphere, of which it forms nearly four-fifths, or, more exactly, 79 per cent. It is there mixed, but not combined with oxygen; and when the latter gas is removed, by introducing into a bottle of air some substance for which the former has an affinity, the nitrogen is left in a state of purity. It is a transparent gas, which is incombustible and extinguishes flame. It is a singularly inert substance, and is incapable of directly entering into union with any other element except oxygen, and with that it combines with the greatest difficulty, and only by the action of the electric spark—a peculiarity which has very important bearings on many points we shall afterwards have to discuss. Nitrogen is found in plants to the extent of from 1 to 4 per cent.

Nitric Acid.—This, the most important compound of nitrogen and oxygen, can be produced by sending a current of electric sparks through a mixture of its constituents, but in this way it can be obtained only in extremely small quantity. It is much more abundantly produced when organic matters are decomposed with free access of air, in which case the greater proportion of their nitrogen combines with the atmospheric oxygen. This process, which is known by the name of nitrification, is greatly promoted by the presence of lime or some other substance, with which the nitric acid may combine in proportion as it is formed. It takes place, to a great extent, in the soil in India and other hot climates; and our chief supplies of saltpetre, or nitrate of potash, are derived from the soil in these countries, where it has been formed in this manner. The same change occurs, though to a much smaller extent, in the soil in temperate climates.

Ammonia is a compound of nitrogen and hydrogen, but it cannot be formed by the direct union of these gases. It is a product of the decomposition of organic substances containing nitrogen, and is produced when they are distilled at a high temperature, or allowed to putrefy out of contact of the air. In its pure state it is a transparent and colourless gas, having a peculiar pungent smell, and highly soluble in water. It is an alkali resembling potash and soda, and, like these substances, unites with the acids and forms salts, of which the sulphate and muriate are the most familiar. In these salts it is fixed, and does not escape from them unless they be mixed with lime, or some other substance possessing a more powerful affinity for the acid with which it is united.

Oxygen is one of the most widely distributed of all the elements, and, owing to its powerful affinities, is the most important agent in almost all natural changes. It is found in the air, of which it forms 21 per cent, and in combination with hydrogen, and almost all the other chemical elements. In the pure state it possesses very remarkable properties. All substances burn in it with greater brilliancy than they do in atmospheric air, and its affinity for most of the elements is extremely powerful. When diluted with nitrogen, it supports the respiration of animals; but in the pure state it proves fatal after the lapse of an hour or two. It is found in plants, in quantities varying from 30 to 36 per cent.

It is worthy of observation, that of the four organic elements, carbon only is fixed, and the other three are gases; and likewise, when any two of them unite, their compound is either a gaseous or a volatile substance. The charring of organic substances, which is one of their most characteristic properties, and constantly made use of by chemists as a distinctive reaction, is due to this peculiarity; for when they are heated, a simpler arrangement of their particles takes place, the hydrogen, nitrogen, and oxygen unite among themselves, and carry off a small quantity of carbon, while the remainder is left behind in the form of charcoal, and is only consumed when access of the external air is permitted.

Now, in order that a plant may grow, its four organic constituents must be absorbed by it, and that this absorption may take place, it is essential that they be presented to it in suitable forms. A seed may be planted in pure carbon, and supplied with unlimited quantities of hydrogen, nitrogen, oxygen, and inorganic substances, and it will not germinate; and a plant, when placed in similar circumstances, shows no disposition to increase, but rapidly languishes and dies. The obvious inference from these facts is, that these substances cannot be absorbed when in the *elementary* state, but that it is only after they have entered into certain forms of combination that they acquire the property of being readily taken up, and assimilated by the organs of the plant.

It was at one time believed that many different compounds of these elements might be absorbed and elaborated, but later and more accurate experiments have reduced the number to four—namely, carbonic acid, water, ammonia, and nitric acid.

The first supplies carbon, the second hydrogen, the two last nitrogen, while all of them, with the exception of ammonia, may supply the plant with oxygen as well as with that element of which it is the particular source.

There are only two sources from which these substances can be obtained by the plant, viz. the atmosphere and the soil, and it is necessary that we should here consider the mode in which they may be obtained from each.

The Atmosphere as a source of the Organic Constituents of Plants.—Atmospheric air consists of a mixture of nitrogen and oxygen gases, watery vapour, carbonic acid, ammonia, and nitric acid. The two first are the largest constituents, and the others, though equally essential, are present in small, and some of them in extremely minute quantity. When deprived of moisture and its minor constituents, 100 volumes of air are found to contain 21 of oxygen and 79 of nitrogen.

Although these gases are not chemically combined in the air, but only mechanically mixed, their proportion is exceedingly uniform, for analyses completely corresponding with these numbers have been made by Humboldt, Gay-Lussac, and Dumas at Paris, by Saussure at Geneva, and by Lewy at Copenhagen; and similar results have also been obtained from air collected by Gay-Lussac during his ascent in a balloon at the height of 21,430 feet, and by Humboldt on the mountain of Antisano in South America at a height of 16,640 feet.

In short, under all circumstances, and in all places, the relation subsisting between the oxygen and nitrogen is constant; and though, no doubt, many local circumstances exist which may tend to modify their proportions, these are so slow and partial in their operations, and so counterbalanced by others acting in an opposite direction, as to retain a uniform proportion between the main constituents of the atmosphere, and to prevent the undue accumulation of one or other of them at any one point.

No such uniformity exists in the proportion of the minor constituents. The variation in the quantity of watery vapour is a familiar fact, the difference between a dry and moist atmosphere being known to the most careless observer, and the proportions of the other constituents are also liable to considerable variations.

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Mammalian Embryology

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Mammal

Mammals (formally Mammalia) are a class of vertebrate, air-breathing animals whose females are characterized by the possession of mammary glands while both males and females are characterized by sweat glands, hair and/or fur, three middle ear bones used in hearing, and a neocortex region in the brain. Mammals are divided into three main infraclass taxa depending how they are born. These taxa are: monotremes, marsupials and placentals. Except for the five species of monotremes (which lay eggs), all mammal species give birth to live young. Most mammals also possess specialized teeth, and the largest group of mammals, the placentals, use a placenta during gestation. The mammalian brain regulates endothermic and circulatory systems, including a four-chambered heart.

There are approximately 5,400 species of mammals, distributed in about 1,200 genera, 153 families, and 29 orders (though this varies by classification scheme). Mammals range in size from the 30–40 millimeter (1- to 1.5-inch) Bumblebee Bat to the 33-meter (108-foot) Blue Whale. Mammals are divided into two subclasses: the Prototheria, which includes the oviparous monotremes, and the Theria, which includes the placentals and live-bearing marsupials. Most mammals, including the six largest orders, belong to the placental group. The three largest orders, in descending order, are Rodentia (mice, rats, porcupines, beavers, capybaras, and other gnawing mammals), Chiroptera (bats), and Soricomorpha (shrews, moles and solenodons). The next three largest orders include the Carnivora (dogs, cats, weasels, bears, seals, and their relatives), the Cetartiodactyla (including the even-toed hoofed mammals and the whales) and the Primates to which the human species belongs.

The relative size of these latter three orders differs according to the classification scheme and definitions used by various authors. Phylogenetically, Mammalia is defined as all descendants of the most recent common ancestor of monotremes (e.g., echidnas and platypuses) and therian mammals (marsupials and placentals). This means that some extinct groups of "mammals" are not members of the crown group Mammalia, even though most of them have all the characteristics that traditionally would have classified them as mammals. These "mammals" are now usually placed in the unranked clade Mammaliaformes.

The mammalian line of descent diverged from an amniote line at the end of the Carboniferous period. One line of amniotes would lead to reptiles, while the other would lead to synapsids. According to cladistics, mammals are a sub-group of synapsids. Although they were preceded by many diverse groups of non-mammalian synapsids (sometimes misleadingly referred to as mammal-like reptiles), the first true mammals appeared in the Triassic period. Modern mammalian orders appeared in the Palaeocene and Eocene epochs of the Palaeogene period.

Distinguishing Features

Living mammal species can be identified by the presence of sweat glands, including those that are specialized to produce milk. However, other features are required when classifying fossils, since soft tissue glands and some other features are not visible in fossils. Paleontologists use a distinguishing feature that is shared by all living mammals (including monotremes), but is not present in any of the early Triassic synapsids: mammals use two bones for hearing that were used for eating by their ancestors.

The earliest synapsids had a jaw joint composed of the articular (a small bone at the back of the lower jaw) and the quadrate (a small bone at the back of the upper jaw). Most reptiles including lizards, crocodilians, dinosaurs (and their descendants the birds) use this system, as did non-mammalian synapsids such as therapsids. Mammals have a different jaw joint, however, composed only of the dentary (the lower jaw bone which carries the teeth) and the squamosal (another small skull bone). In mammals the quadrate and articular bones have become the incus and malleus bones in the middle ear.

Mammals also have a double occipital condyle: they have two knobs at the base of the skull which fit into the topmost neck vertebra, and other vertebrates have a single occipital condyle. Paleontologists use only the jaw joint and middle ear as criteria for identifying fossil mammals, since it would be confusing if they found a fossil that had one feature, but not the other.

Anatomy and Morphology

Skeletal System

The majority of mammals have seven cervical vertebrae (bones in the neck); this includes bats, giraffes, whales, and humans. The few exceptions include the manatee and the two-toed sloth, which have only six cervical vertebrae, and the three-toed sloth with nine cervical vertebrae.

Respiratory System

The lungs of mammals have a spongy texture and are honeycombed with epithelium having a much larger surface area in total than the outer surface area of the lung itself. The lungs of humans are typical of this type of lung.

Breathing is largely driven by the muscular diaphragm which divides the thorax from the abdominal cavity, forming a dome with its convexity towards the thorax. Contraction of the diaphragm flattens the dome increasing the volume of the cavity in which the lung is enclosed. Air enters through the oral and nasal cavities; it flows through the larynx, trachea and bronchi and expands the alveoli. Relaxation of the diaphragm has the opposite effect, passively recoiling during normal breathing.

During exercise, the abdominal wall contracts, increasing visceral pressure on the diaphragm, thus forcing the air out more quickly and forcefully. The rib cage itself also is able to expand and contract the thoracic cavity to some degree, through the action of other respiratory and accessory respiratory muscles. As a result, air is sucked into or expelled out of the lungs, always moving down its pressure gradient. This type of lung is known as a bellows lung as it resembles a blacksmith's bellows.

Nervous System

All mammalian brains possess a neocortex, a brain region that is unique to mammals.

Integumentary System

The integumentary system is made up of three layers: the outermost epidermis, the dermis, and the hypodermis.The epidermis is typically ten to thirty cells thick, its main function being to provide a waterproof layer. Its outermost cells are constantly lost; its bottommost cells are constantly dividing and pushing upward. The middle layer, the dermis, is fifteen to forty times thicker than the epidermis. The dermis is made up of many components such as bony structures and blood vessels. The hypodermis is made up of adipose tissue. Its job is to store lipids, and to provide cushioning and insulation. The thickness of this layer varies widely from species to species.

Although mammals and other animals have cilia that superficially may resemble it, no other animals except mammals have hair. It is a definitive characteristic of the class. Some mammals have very little, but nonetheless, careful examination reveals the characteristic, often in obscure parts of their bodies. None are known to have hair that naturally is blue or green in colour although some cetaceans, along with the mandrills appear to have shades of blue skin. Many mammals are indicated as having blue hair or fur, but in all known cases, it has been found to be a shade of gray. The two-toed sloth and the polar bear may seem to have green fur, but this colour is caused by algae growths.

Reproductive System

Most mammals give birth to live young (vivipary), but a few, namely the monotremes, lay eggs. The platypus and the echidna present a particular sex determination system that is different from other vertebrates. Certain glands of mammals known as mammary glands are specialized to produce milk, a liquid used by newborns as their primary source of nutrition. The monotremes branched early from other mammals and do not have the nipples seen in most mammals, but they do have mammary glands.

Viviparous mammals are classified into the subclass Theria and are divided into two infraclasses: Metatheria (of which only the Marsupialia survive), and Eutheria. Marsupialia, or marsupials, have short gestation periods and give birth to undeveloped young which are contained within a pouch-like sac (marsupium) located in front of the mothers' abdomen. Eutherians, commonly known as placentals, are mammals that give birth to complete and fully developed young. This is usually characterized by long gestation periods. The majority of mammal species are classified as eutherians.

Physiology

Endothermy

Nearly all mammals are endothermic ("warm-blooded"). Most mammals also have hair to help keep them warm. Like birds, mammals can forage or hunt in cold weather and climates where reptiles and large insects cannot. Endothermy requires plenty of food energy, so pound for pound mammals eat more food than reptiles. Small insectivorous mammals eat prodigious amounts for their size. A rare exception, the naked mole rat produces little metabolic heat, so it is considered an operational poikilotherm. Birds are also endothermic, so endothermy is not a defining mammalian feature.

Intelligence

In intelligent mammals, such as primates, the cerebrum is larger relative to the rest of the brain. Intelligence itself is not easy to define, but indications of intelligence include the ability to learn, matched with behavioral flexibility. Rats, for example, are considered to be highly intelligent as they can learn and perform new tasks, an ability that may be important when they first colonize a fresh habitat. In some mammals, food gathering appears to be related to intelligence: a deer feeding on plants has a brain smaller than a cat, who must think to outwit its prey.

Earliest Appearances of Features

Hadrocodium, whose fossils date from the early Jurassic (approx. 195 million years ago), provides the first clear evidence of fully mammalian jaw joints. It has been suggested that the original function of lactation (milk production) was to keep eggs moist. Much of the argument is based on monotremes (egg-laying mammals):

The earliest clear evidence of hair or fur is in fossils of *Castorocauda*, from 164M years ago in the mid Jurassic. From 1955 onwards some scientists have interpreted the foramina (passages) in the maxillae (upper jaws) and premaxillae (small bones in front of the maxillae) of cynodonts as channels which supplied blood vessels and nerves to vibrissae (whiskers), and suggested that this was evidence of hair or fur. But foramina do not necessarily show that an animal had vibrissae – for example the modern lizard *Tupinambis* has foramina which are almost identical to those found in the non-mammalian cynodont Thrinaxodon. The evolution of erect limbs in mammals is incomplete — living and fossil monotremes have sprawling limbs. In fact some scientists think that the parasagittal (non-sprawling) limb posture is a synapomorphy (distinguishing characteristic) of the Boreosphenida, a group which contains the Theria and therefore includes the last common ancestor of modern marsupial and placentals – and therefore that all earlier mammals had sprawling limbs. *Sinodelphys* (the earliest known marsupial) and *Eomaia* (the earliest known eutherian) lived about 125M years ago, so erect limbs must have evolved before then.

It is currently very difficult to be confident when endothermy first appeared in the evolution of mammals. Modern monotremes have a lower body temperature and more variable metabolic rate than marsupials and placentals. So the main question is when a monotreme-like metabolism evolved in mammals. The evidence found so far suggests Triassic cynodonts may have had fairly high metabolic rates, but is not conclusive. In particular it is difficult to see how small animals can maintain a high and body temperature without fur.

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The History of Forensic Entomology

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In an experiment famous as much for its demonstration of scientific method as for its contribution to entomology, Francesco L. Redi (1668) studied rotting meat that was either exposed to or protected from flies. From his analysis of subsequent blow fly infestation, he refuted the hypothesis of the "spontaneous generation" of life. Up to that time, it was generally believed that under the right conditions maggots came from rotten meat. Later, Bergeret (1855), near Paris, France, was the first westerner to use insects as forensic indicators. The body of a baby was found behind the plaster mantle in a house, and an investigation was begun. Bergeret determined that the assemblage of insects associated with the corpse pointed to a state of decay that dated back several years; consequently, the question of guilt was thrown upon the earlier occupants of the house, and not upon the current ones.

Bergeret's methods and materials were quite similar to one of the main medicocriminal entomological techniques still in use today; that is, the successive colonization of a corpse by a predictable succession of arthropod species. Between 1883 and 1898, J. P. Megnin in France published a series of articles dealing with medicocriminal entomology. The most famous of these, *La Faune des Cadares*, served in large part to make the medical and legal professions aware that entomological data could prove useful in forensic investigations.

Although entomologists are most familiar with the references cited above, medicocriminal matters in the Far East predate these considerably. In 1235 A.D., Sung Tz'u, a Chinese "death investigator," wrote a book entitled *The Washing Away of Wrongs* in which forensic science as known at that time was detailed. In this text, what was probably the first actual medicocriminal entomology case was recounted. A murder by slashing occurred in a Chinese village, and the local death investigator was deputized to solve the crime. After some fruitless questioning, the investigator had all villagers bring their sickles to one spot and lay them out before the crowd.

Flies were attracted to one of the sickles, probably because of invisible remnants of tissue still adhering to it, and the owner subsequently broke down and confessed to the crime. In other portions of the text, Sung Tz'u demonstrated knowledge of blow fly activity on bodies relative to those orifices infested, the time of such infestation, and the effect of trauma on attractiveness of tissue to such insects.

Any analytical system is as reliable as is the data upon which it is founded, and forensic entomology is no exception. Because accurate identification of necrophilous arthropods is of paramount importance, few repeatable results could be obtained before adequate taxonomic work had been accomplished on the invertebrates (the insects and related animals) in question. Taxonomy and systematics comprise the science describing, classifying, and proposing evolutionary relationships of the various forms of life.

Although many synanthropic (strongly associated with human activity) flies (such as Drosophila, Musca, Muscina, Ophyra, Stomoxys, and others) are not encountered frequently in typical forensic investigations, other species assume great importance. Carrion (dead tissue) feeding blow flies (Calliphoridae) and flesh flies (Sarcophagidae) are those most useful in death investigations. Aldrich's (1916) monograph on the Sarcophagidae made use of distinctive male genitalia, thereby enabling entomologists to identify adult male specimens from this important family.

This concept involved the so called "lock-and-key" arrangement in many insects that facilitates reproductive isolation between species. The male copulatory organs of each kind (species) of higher flies are composed of unique, complex structures that are used as key characters to enable specific determination. This adaptation has been applied with equal success to the forensically important blow flies.

Twenty years later, Knipling (1936) published descriptions and keys to many common early (first instar) maggots of flesh flies. Although considerable work had been done on the blow fly fauna of North America (for instance, Knipling 1939), Hall's 1948 monograph, *The Blowflies of North America,* made possible the accurate identification of adults and mature larvae of most species of this family as well.

Although very few new (that is, previously unrecognized) North American calliphorid species have been described recently, efforts have been devoted to accumulate improved distributional information (Hall and Townsend 1977, Hall 1979, Goddard and Lago 1983). More research is needed on accurate identification of the critical larval and pupal stages (those most frequently collected in death investigations). At present, first instar blow fly larvae (the stage that hatches directly from the egg) generally are not identifiable to species, and second instars (the next maggot stage) can be identified accurately only on occasion.

The situation is somewhat better with respect to third instar or prepupal larvae (the largest maggot stage, and that most commonly observed), but only if such specimens are preserved properly. Even so, a significant number of indigenous blow flies cannot be identified at present as immatures. This is currently an area of active research, and to this end the relatively new technique of scanning electron microscopy is being applied.

Because of the medicocriminal requirement for reliable data on rates of larval development, considerable effort has been expended to measure such intervals. Anecdotal information on blow flies contained in earlier works was largely supplanted by Hall's (1948) rearing data, and the latter has been refined for some forensically important species to degree hour status. Because insects are coldblooded animals, their rate of development is more or less dependent on ambient temperature. Research has shown that for each species there generally is a threshold temperature below which no development takes place.

As temperature rises above this threshold, a certain amount of time is required for the insect to attain defined stages of development (for instance, from the newlylaid egg through the second instar maggot). Because this heat is accumulated as "thermal units," it can be calibrated and described as "degreedays" or "degreehours," depending on the accuracy of temperature readings and time period involved.

However, most laboratory rearings (upon which the degreehour data are developed) have been done at constant temperature, so additional research will be necessary to establish correlations between these data, typical fluctuating field temperatures (warmer during the day and cooler at night), and the average daily measurements frequently reported from weather stations. Retrospective weather records from the nearest weather recording station (such as an airport) are those most often used in medicocriminal evaluations.

Access to the scientific literature pertaining directly to medicocriminal entomology has been facilitated by two recent bibliographies. An initial guide to entomological involvement in forensic pathology, plus a selected bibliography, was provided by Meek et al. (1983). A bibliography of all publications dealing wholly or in large part with medicocriminal entomology worldwide was compiled by Vincent et al. (1985). The latter paper contained 329 references and was current through 1983; therefore, the actual body of literature pertaining to this subdiscipline of forensic entomology is not large when compared to many other biological or legal subjects. The first textbook devoted to forensic entomology was published in 1986: *A Manual of Forensic Entomology.* This is an excellent reference for the entomologist, and it brings together in one place all the salient information contained in the literature on this subject.

A procedural guide, *Entomology and Death*, was published in 1990 and is intended for crime scene investigators and other forensic specialists.

Scale Insects

Scale insects are a diverse group of insects in the order Hemiptera. There are about 6,000 species of scale insects in 21 families worldwide. About 1,000 species occur in North America. The three most common families of scale insects are the armoured scale, the soft scale, and the mealybugs. Most of the pest species belong to one of these three families.

Armoured Scale Insects: Armoured scales are the smallest of scale insects, ranging in size from 1 to 3 millimetres. The body of the scale insect is protected by a cover (the armour) made from wax secreted by the insect and cast skins (exuviae) of previous growth stages. One must remove the hardened wax cover to expose the body of the insect. The exposed body usually is yellow or orange, but may have a pink or red colour to it. This cover also protects the eggs laid by the female. Armoured scale insect covers vary from circular to elongate or oyster shell-shaped. Male and female covers may differ in size and shape for the same species. The cover of the female is generally largest. Boisduval scale and fern scale are common armoured scale insects attacking flowers and foliage plants.

Most armoured scale insects reproduce sexually. The eggs hatch beneath the protective scale cover and the first instars, commonly called “crawlers”, migrate to the new growth to settle and feed. Armoured scale females lose their legs at the first molt and are sessile for the rest of their lives. Females develop through three instars and males develop through five. Armoured scales may overwinter as eggs, nymphs, or adult females. Adult males are usually present about two weeks in each generation. Some armoured scales have four generations per year.

Soft Scale Insects: Soft scales differ from armoured scales in that they do not secrete a waxy covering that is separate from the body.

If wax is present, it adheres tightly to the body of the female and cannot be easily separated from it. Most soft scales produce a thin, glassy wax that does not obscure the colour or form of the female soft scale.

Soft scales are fairly large (2 to 6 millimetres long) and can be distinguished by their larger size, round or oval body outline, and convex or hemispherical profile. Soft scale females vary from flat to almost spherical. Often different host plants will alter the body form of a single species so much that taxonomists have described the different forms as separate species. If one turns the adult soft scale over, legs, antennae and thread-like mouthparts are readily visible with the aid of a microscope. Three common soft scales found in greenhouses and interior plantscapes are the brown soft scale, hemispherical scale, and tessellated scale. Soft scales may reproduce sexually or parthenogenically and every female may be capable of producing progeny without fertilization. Tremendous populations can develop during a single growing season. Most outdoor species have one generation per year. Females either lay eggs or give live birth, depending on the species.

There are three instars in the females and five instars in the males. In warmer climates and in greenhouses, species with multiple generations may have all stages present simultaneously throughout the year.

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Biological Control and Holistic Plant-health Care in Agriculture

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In considering the contributions of biological pest control to a sustainable agriculture, it may be useful first to examine briefly some of the advantages and disadvantages of each of the major methods by which pests can be controlled. The major methods of pest control can be grouped into three categories of 1) physical control, 2) chemical control, and 3) biological control. These broad categories, in turn, can be combined into integrated pest management (IPM), integrated crop and pest management (ICPM), or, as will be used in this article, holistic plant-health care or simply plant-health caret The equivalent for livestock is integrated livestock management or animal health care.

Physical control includes tillage to control weeds, open-field burning to control pests (Hardison, 1976), solar heating the soil beneath clear plastic tarp, elimination of pathogens from milk or rooting media by mild heat-treatment , the production of pathogen-free plants from tissue culture started with clean meristem or shoot tips , and the physical separation of crop from a potential pest attack by choice of planting date. The production of pathogen-free plants from tissue culture is nonpolluting and, along with indexing of seeds, is the best or only acceptable method to eliminate some viral and bacterial pathogens so that the planting material can be certified as pathogen-free. On the other hand, tillage is energy-expensive (Phillips et al., 1980) and contributes to soil erosion; the trend in the United States is therefore toward less tillage to conserve energy, reduce soil erosion, and make U. S. agriculture more sustainable. Open-field burning contributes to air pollution and may, over the long term, have a negative effect on the organic master content of soil; the tendency is therefore toward reduced use of burning, and legislation has been introduced in some states to regulate or even ban open-field burning. Solar-heating the soil involves the capture of incoming solar radiation beneath clear plastic sheeting placed on the soil surface. It is a safe method by which plant-parasitic nematodes, soilborne fungal pathogens, soil-inhabiting insect pests, and some weed seeds can be eliminated by heat treatment of soil in gardens, vegetable fields, and orchards , but is usually not economical except for high-value crops and in areas of abundant sunshine.

Chemical control is used in this report to mean control of pests with chemical pesticides. The problems of chemical pesticides have been reviewed amply and need not be restated here. While some pesticides must be abandoned because of their unacceptable nontarget effect, there will always be a need in agriculture for safe and selective chemicals to limit the effects of pests. More significantly, it is becoming increasingly more difficult and expensive to find new kinds of synthetic chemical pesticides. The chemical pesticide industry has therefore been described as a "maturing industry."

Biological control is the control of one organism by another (Beirner, 1967). This control may be expressed as either a longer population of the pest (DeBach, 1964) or as a restriction or prevention of the severity or incidence of pest damage without regard to the pest population. Biological control depends on knowledge of biological interactions at the ecosystem, organismal, cellular, and molecular levers and often is more complicated to manage compared with physical and chemical methods. Biological control is also likely to be less spectacular than most physical or chemical controls but is usually also more stable and longer lasting. In spite of biological controls having been used in agriculture for centuries, as an industry biological control is still in its infancy.

Biological control is now being considered for an increasing number of crops and managed ecosystems as the primary method of pest control. One reason for its growing popularity is its record of safety during the past 100 years considered as the era of modem biological control. No microorganism or beneficial insect deliberately introduced or manipulated for biological control purposes has, itself, become a pest so far as can be determined, and there is no evidence so far of measurable or even negligible negative effects of biocontrol agents on the environment. Another reason for considering biological control over other methods is untapped potential; biological control is underused, under exploited, underestimated and often untried and therefore unproven. The new tools of recombinant DNA technology, mathematical modelling, and computer technology combined with a continuation of the more classical approaches such as importation and release of natural enemies and improved germplasm, breeding, and field testing should quickly move biocontrol research and technology into a new era.

Biological control describes the normal state of affairs in natural undisturbed ecosystems, where populations of organisms exist in a dynamic equilibrium and species or individuals unable to compete or to find an ecological niche are replaced by those that can. With sufficient knowledge, it becomes possible to manipulate this equilibrium so as to favour some organisms more than others. Thus be gins agriculture , silviculture , gardening, and other similar activities that favour a few desirable plant or animal species, or subsets of species (cultivars, breeds, strains), that otherwise could not succeed and might even become extinct.

This chapter is focused on biological control of pests and diseases of plants important in farmland, orchards, and other agroecosystems. Many of the examples discussed involve the control of diseases of wheat (Cook, 1986c), but the concepts presented are just as applicable to pest and disease control on other crops and in other managed ecosystems, including urban and recreational areas.

This chapter also introduces some principles of holistic plant-health care, which involves extensive use of biological control integrated with physical and chemical treatments and pest controls as appropriate and compatible with the goals of making agriculture more sustainable.

Biological Control as a Concept

Biological control was discovered by trial and error and then practiced in agriculture long before the term itself came into use. One example is the ancient practice of not growing the same crop species in the same field more frequently than every second or third year or even longer. Such crop rotation allows time for the pest or pathogen population in soil to decrease below some economic threshold because of the predatory, competitive, and other antagonistic effects imposed by the associated microflora and fauna. In other words, crop rotation allows time for the natural soil microbiota to sanitize the soil, especially with regard to the more specialized plant parasites and insect pests that are highly dependent on their host crop to maintain their populations.

The era of modern biological control, involving the deliberate transfer and introduction of natural enemies of insect pests, was launched 100 years ego with the highly successful introduction of the vadalia beetle from Australia to California in 1888 to control the cottony cushion scale of citrus. In 1914, the German plant pathologist C. F. von Tubuef wrote a somewhat speculative article entitled "Biologische Bekampfung von Pilzkrankheiten der Pflanzen." This is apparently the first reference in the scientific literature to the term "Biologische Bekampfung" or "biological control" (Baker, 1987). In 1916, L. O. Howard referred to control of the cottony-cushion scale insect by the vadalia beetle as a "biological method" and in 1919, H. S. Smith called it biological control.

About 80 years ago, a gene for resistance in wheat to wheat stem rust caused by Puccinia graminis f. sp. tritici was successfully transferred for the first time by crossing a rust-resistant with a rust-susceptible wheat plant (Biffen, 1906). Thus began the practice of introducing genes for resistance to pests, first by conventional methods and now expanded to include genetic engineering by use of recombinant DNA technology. Lupton (1984) gave emphasis to this approach in his presidential address to the Association of Applied Biologists in Great Britain, entitled, "Biological Control: The Plant Breeder's Objective." Moreover, the boundaries that once existed between these two approaches to biological control—transfer of whole organisms and transfer of genes—are beginning to disappear because of the tools of recombinant DNA technology. For example, a gene for production of an endotoxin by a strain of Bacillus thuringiensis, lethal to certain insect pests of crops, has now been transferred by recombinant DNA technology to tobacco and tomato and shown to function in these plants as genes for resistance to the toxin-sensitive insect pests of these plants.

It is commonly argued that biological control as a concept should exclude host plant resistance to pests and diseases achieved by introduction of genes through plant breeding (R. Baker, 1985). Such a definition puts this science into the awkward position that the use of a plant pathogen with certain genes for virulence to maintain a population of susceptible weed plants at or below some economic threshold would qualify as biological control, but the converse, the maintenance of a pathogen population at or below some economic threshold by deployment of certain genes for resistance in the crop plants, would not qualify as biological control.

As another incongruity, the bt gene for production of the insect toxin expressed in the insect pathogen Bacillus thuringiensis would qualify as biological control but the same gene expressed in plants, as a gene for resistance to insect pests, would not qualify as an example of biological control. Such a narrow definition is artificial and scientifically indefensible. Perhaps Lupton (1984) has stated it best: "accelerating or diverting evolutionary processes in order to obtain genotypes adapted to [man's] needs are a most important example of the application of biological control to agricultural and horticultural crops."

DeBach (1964) defined biological control as "the action of parasites, predators, or pathogens in maintaining another organism's population density at a longer average than would occur in their absence." This definition covers some highly successful biological controls of insect pests with natural enemies, but it does not accommodate some other highly successful controls accepted in other disciplines as examples of biological control.

For example, citrus tristeza virus is controlled in Brazil by inoculating the citrus trees with a mild virus, which then protects the trees against the more severe strains (Costa and Muller, 1980). "Cross protection" was first shown by H. H. McKinney in 1929 to have potential for biological control of plant viruses. Plant pathologists refer to cross protection for control of plant viruses as biological control.

The many biological controls that fall outside the narrow definition have been variously labelled as "biological methods of control," "biological forms of control," and "biological pest suppression". All of these terms, like H. O. Howard's original "biological method," mean simply "biological control."

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Nanomaterials: Small Miraculous Particles

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In the near future Science and Technology will depend on the materials of very small size called nanomaterials.These materials have potential applications in various fields. Nanotechnology has the potential to revolutionize the methodology for detection of diseases and to play a major role in management of various problems. It is obvious that nanomaterials will take part in each and every area of new product development and their impact will be such that many of the limitations of current approaches will be overcome.

Key words: nanomaterials, carbon nanotubes, nano robots, nano solar cells.

Introduction

Nanoscience and nanotechnology are among the fastest emerging areas of research. Greek word nano means dwarf and one nanometer is equal to one billionth of a meter. Nanomaterials are the substances which have dimensions in the range 1-100 nm. Nanoscience is the study of fundamental principles governing structure, properties, synthesis and applications of particles with at least one dimension roughly ranging from 1 to 100 nanometers. Nanomaterials are classified on the basis of their dimensions. Thin film layers, surfaces and coatings are one dimensional nanomaterials. Two dimensional nanomaterials include nanotubes, nanowires, nanofibres and nanopolymers. Fullerenes, dendrimers and quantum dots are three dimensional nanomaterials.

Properties of Nanomaterials

Nanotechnological advancement in modern time is of great relevance for scientists and researchers as substances in nanoform have unique physical, chemical and biological properties which are needed to be tested. Many of these properties are of great importance as these may be employed for specific uses in different fields e.g.in health care, in technical fields and various others. As a particle decreases in size a greater proportion of atoms are found at surface compared to those inside. Due to a greater surface area per unit mass compared to larger particles, quantum effects begin to dominate the properties of nanomaterial. At the nanoscale most of the fundamental properties of materials and machines depend on their size in a way they do not at any other scale. Metal-nanotechnological substance interaction has been a topic of concern for scientists. In the field of corrosion science, utility of nanotechnological materials is studied by researchers e.g. use of polyaniline fabricated alloys has been studied for prevention against corrosion.

Synthesis of Nanomaterials

The properties of nanoparticles depend not only on their composition but also on their shape and size. In order to create nanoparticles with newer properties and features it is important to synthesize complex nanostructures. The synthesis methodologies of nanomaterials are based on trial and error in nature. Two approaches have been adopted for the synthesis of nanoparticles. Top down approach involves the synthesis of nanostructures starting with largest structures and taking its parts away e.g. ball milling technique. In this, grinding of samples is done down to less than 100nm and can be used, homogenizing and alloying. The other approach i.e. bottom up approach involves the use of atoms or molecules as starting materials e.g. wet process or aerosol process. In this well defined quantities of different ionic solutions are mixed. Insoluble compound is precipitated under controlled temperature and pressure. The precipitate is filtered or spray dried to get the nanoparticles in powdered form. Using this method it has been made possible to produce single and multicomponent materials with controlled levels of doping with many elements. The development of multifunctional nanosystems has helped in achieving multifunctionality.

Applications of Nanotechnology

One dimensional nanomaterials are corrosion resistant, wear and scratch resistant materials. These are often used as self cleaning, dirt repellent, antibacterial, antimicrobial, catalytically active, chemically functionalized and transparency modulated nanomaterials. Nanotechnology can be applied to revolutionize the technique of detecting and treating diseases. Nano robots can score inside the human body seeking out harmful germs and destroying them. Nanorobots are also capable of delivering drugs at certain targets enhancing the effectiveness of drugs without touching the surrounding healthy tissues. This technology of treatment is far more superior than chemotherapy used in the treatment of diseases like cancer where safety of healthy cells in the vicinity of cancerous cells is not certain.

Nanomaterials can be used to control pollution also. Nanotechnology involving pumping of nanomaterials into the ground is aimed at converting hazardous chemicals present in ground water into harmless products. In today's world there is an increasing awareness about water conservation and desalination of water at a major scale. Reverse osmosis technology is applied to remove salinity of water .Here carbon nanotubes (CNT) are very useful as they are hydrophobic. Due to small diameter of CNT, charged species are excluded from permeation across these. Water molecules can be separated from salt with a greater efficiency by forcing them through networks of carbon nanotubes which require lower pressures than conventional reverse osmosis.

Advancement in nanoscale science and technology has resulted in solutions of most of the problems concerning water treatment using nanosorbents , nanocatalyst, nanoparticle bioactive nanoparticles, nanostructured catalytic membranes, nanoparticle enhanced filtration e.t.c. Catalysts made from nanoparticles offer a greater surface area for the interaction with the reacting chemicals than catalysts made from larger particles allowing more chemicals to interact with the catalyst simultaneously.

By applying nanotechnology, light weight solar cells can be made. Travelling in space can be less costly if solar cells are used instead of rocket fuel. Researches are going on to use nanotechnology to develop nano solar cells that would be energy intensive and less expensive. Plasmonic nanostructures have been reported to enhance the efficiency of some solar cells. It will not be a surprise if a solar power station is built in space in near future which will be capable of harnessing solar energy and transferring it to earth everyday with a nine times more efficiency than the solar cells on earth.

From a long time scientists have been involved in development of a clinlcally and economically viable ,ultra sensitive sensor system. Nanotechnology is applied in developing improved, rapid, ultra-sensitive, multiplexed and point of care detection systems. Surface modification of nanoparticle components with suitable organic and inorganic reagents is required for the generation of multifunctional nanoplatforms. For nanotechnologists, the exact nature of interaction between nanoparticles and surface modifying reagents is the subject of research.

The research in inorganic chemistry deals with dense and porous films with thickness ranging from 1nm to several microns.

The films are prepared by various techniques like ALD, CVD, PVD, sol-gel and electrodeposition.

Carbon Nanotubes : Wonderful Material

There are various potential applications of carbon nanotubes. Water proof and tear resistant fabrics can be made by using carbon nanotubes (CNT). Combat jackets use CNT fibres to stop bullets and to monitor the condition of the wearer. CNT is useful in increasing tensile strength of concrete and to halt crack propagation. Elastic modulus of polyethylene polymer is increased by 30% on adding CNT to it. Stronger and lighter tennis rackets, bicycle parts, golf balls, golf clubs and base ball parts are made by using CNT.Investigations are going on possible use of CNT in space elevator technology where tensile strength of more than about 70Gpa is required.

CNT are ideal for synthetic muscles due to high contraction/extension ratio. Fibres produced with polyvinyl alcohol have high tensile strength which require 600J/g to break. CNT may be used to replace steel in suspension and other bridges. Due to high strength/weight ratio CNTs are used in fly wheels to provide high rotational speeds. Single walled carbon nanotube aligned in parallel can be elastically stretched for an energy density greater than that of current lithium ion batteries, with many additional advantages. Thin layers of buckypaper can significantly improve fire resistance due to the efficient reflection of heat by the dense ,compact layer of CNT or carbon fibres.

CNT can be used to produce nanowires of other elements/molecules such as gold or zinc oxide. These nanowires in turn can be used to cast nanotubes of other chemicals such as gallium nitride.These have very different properties from CNTs. For example gallium nitride nanotubes are hydrophilic while CNTs are hydrophobic, making them useful for applications in Organic Chemistry. Scientists have been developing CNT films and Nano Buds to replace indium tin oxide (ITO) in LCDs, touch screens and photovoltaic devices. Nanotube films show promise for use in displays for computers, cell phones, personal digital assistants and automated teller machines. A nanotube formed by joining two nanotubes of different diameters end to end can act as a diode, suggesting the possibility of constructing computer circuits entirely of nanotubes. Additional advantage is that CNT have good transmission properties thus heat is dissipated potentially from computer chips. Researches have shown that carbon nanotubes are suitable scaffold material for osteoblast proliferation and bone formation.

CNT membranes can filter carbon dioxide from power plant emissions thus are useful in manufacture of pollution control equipments. CNT have potential to store between 4.2 - 65 % hydrogen by weight.

Conclusion

Nanoscience and Nanotechnology are emerging areas of study having immense scope in different fields. Synthesis of nanomaterials and their application for the benefit of mankind is going to replace the whole items we use in our daily life. Future world will be a place with most of the materials made of nanoparticles and most of the appliances and instruments based on nanoscience & nanotechnology.

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Factors Influencing in Learning of Mathematics at Secondary Level

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Abstract

Learning Mathematics concept is necessary and should be required for all students. As Mathematics become part of daily life, the need for proper learning has become extremely essential. Because of this increase in needs, demands have also need placed in secondary school to education students and make them "Mathematically equipped". But it is a matter of concern that millions all over the globe have developed dislike and fear for learning Mathematics. The result of different Boards final examination indicate that the failure rate in Mathematics at secondary level in very high. Having realized the importance of Mathematics in all areas of life, the present study is made if order to find out some major factors that influences in learning of Mathematics at secondary level and to gain a deeper understanding of the way in which these factors play role.

However, the factors like classroom environment, teacher's role, teacher's professional experience and job satisfaction, parent's involvements, students' attitudes, teaching methodology, teacher- student ratio, teaching aid etc. are responsible for learning Mathematics at secondary level. Teacher's encouragement found to be a major influencing factor in creating a positive attitude towards learning of the subject Mathematics. It has also been observed that 99.5% of students of secondary level feel that teacher's encouragement and good relation with students help student to learn the subject Mathematics and to achieve good result.

Key Words: Factors, learning, Secondary Level.

Introduction

Secondary school Mathematics education is very different from elementary level not only in the depth of the content but also in attitudes of students towards learning of Mathematics. Study of Mathematics at secondary level is the foundation stage of higher education. The education commission of India (1964-65) has recommended Mathematics as a compulsory subject for the students of secondary level.

It is undoubtedly true that a strong Mathematical background is necessary for many career and job opportunities in increasingly technological society of present time. But unfortunately many potential secondary school students restrict their Mathematics learning and career option by the rapid growth of negative Mathematics learning attitude. Even some of those who continue to study Mathematics at secondary level and onwards do not demonstrate satisfactory achievement in Mathematics. It has been observed that even today percentage of failure in Mathematics is quite high compare to other subjects at secondary level. In Indian educational system failure in secondary level is very common feature. In Assam also a substantial portion of students fail in HSLC examination every year. In 1985 the percentage of failure was 71% and after 15 years in 2010 though it is decreased to 36.79% but not satisfactory.

The results of different Board of Secondary Education of our Country indicate that s large number of students always fail in Mathematics under each and every Board. Tewari (1970) pointed out that 50% of the students fail in Matriculation examination due to failure in Mathematics. It has been affirmed by most of the higher secondary and advisory Boards of our Country like CABE (Central Advisory Board of Education), DEPSE (1964), NCERT (1964), GCPI (1964) etc that Mathematics is the killer subject in HSLC examination. It has been observed that students face different types of difficulties in learning Mathematics which causes failure in the subject.

The interactions of a large number of socio- economic as well as academic environmental factor influences the students and create difficulties in learning Mathematics. Poor learning of Mathematics at secondary level not only results in the students having a low self esteem and poor school performance but also causes significant stress to the parents (Karande and Kulkarni, 2005) Identification of such factories that influences students in learning Mathematics and causes of poor performance in Mathematics in particular and execution of corrective action plan so that the students can perform in Mathematics and enhance their overall percentage of marks up to their full potential in required.

Literature Review

S.L. Jain and G.L. Bured (7) have found in their investigation that low standard of teaching, lack of timely correlation of home work, insufficient periods of teaching Mathematics, lack of appropriate class room etc are the factors responsible for the deficiate learning in Mathematics and low result in secondary Mathematics in Rajasthan.

M.M. Chel (2) in his study suggests greater motivation of the students for learning Mathematics, removal of fear for mathematics and clearer presentation of the subject based on the need of the children.B.S. Kasat (8) has made an in-dept study in Marathi Medium High School students and found that students having poor intelligence, poor study habits, poor numerical ability, poor comprehension and recall ability, lack of help from parents and teacher had no interest in learning Mathematics and failed in S.S.C examination.

D. Hariharan (5) advocated that every Mathematics teacher should assign homework in Mathematics to develop of positive attitude towards learning of Mathematics and better performance in Mathematics of the students.

Method

Population and Sample

The present study consists of secondary school students of Kamrup District of Assam as the population.

The student of class X has selected because this standard is the last standard of secondary school and such acquires as much knowledge as prescribed for him in the High School Syllabus.

Sample has been selected in two phases first phases secondarys in random order and then student and teacher from these school have been considered.

Data Collection

Date was collected primarily through a well defined questionnaire from the sample.

Analysis of Data

A student profile was developed on the bases of information and data collected through survey and interview. Data were analyzed in a very simple way using the statistical tools like Bar diagram, line diagram, Pie diagram etc statistical table etc considering students performance in Mathematics and learning interest of the subject and the impact of factors arises behind it.

Findings and Conclusion

The findings of the present study indicate deeper understanding about the factors influencing in learning mathematics at secondary level. The factors like inadequate Teacher Preparation, Teachers professional experience and job satisfaction, imparting of limited knowledge of Mathematics in the class room, absence of Mathematical approach, study hour, teacher student ratio, teaching aid, teaching methodology, teachers students good relation and teachers encouragement, parents involvement, students attitudes etc having great impact on students in learning Mathematics at secondary level.

However, the present study emphasized on teacher students good relation and teacher encouragement as it has a direct impact on students in creating positive attitude towards learning Mathematics. Less teacher student ratio not only maintains a good class room environment but also helps students in learning mathematics being able to draw a direct attention and importance of a teacher. Professionally trained up teacher with latest technological teaching aid, having practical experience of Mathematical concepts and methods and involvement of parents at all steps as students spent most of the time at home will enable students to be a good learner of mathematics at secondary level onwards.

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Boundary Layer Along a Porous Wall in a Conically Source Flow

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Abstract

In the present analysis an attempt has been made to study the boundary layer of an incompressible fluid in a conically source flow with homogeneous suction at the wall (space source flow). The momentum integral and the K.E. integral equation have been numerically solved step-by-step by the Runga-Kutta method for a few steps and then by Adam's method using a quadrature formula. The point of separation has been determined for solid wall as well as porous wall.

Keywords: Source, Suction, Runga-Kutta method.

Introduction

Waston made an investigation into the existence of similar solution of two dimensional boundary layer equation with suction and found that similar solution exist when the external velocity of the form $U(x)$ - cx^m and suction velocity is proportional to $x^{\frac{m-1}{2}}$.

For a solid wall Rott and Crabtree expounded a method of direct integration by a quadrature formula for the calculation of the momentum thickness for axi-symmetric boundary layers with favourable pressure gradient. Sinha and Chaudhary studied the boundary layer in divergent channel wall with uniform suction. In the present analysis an attempt has been made to study the boundary layer of an incompressible fluid in connically source flow with homogenious suction at the wall. It has been assumed that the permeable wall being from $x = 0$ and velocity at the leading edge of the wall is U_0. This would be possible if a conical wall is placed along the stream lines in a space source flow with the vertex at the source.

The boundary layer thickness at the leading edge will be zero. It is known that in the region of accelerated flow the joint use of the momentum integral and the kinetic energy integral equation with a highly infinite system of velocity profiles. Because of singularity is kinetic energy integral equation at the leading edge ($x = a$), the momentum integral only has been solved by Runga-Kutta method over a short distance with the satisfaction of the wall compatibility conditions.

Then the momentum integral and the Kinetic energy integral equations have been numerically solved step-by-step by the Runga-Kulta method for a few steps and then by Adam's method using a quadrature formula.

The point of separation has been determined for solid wall as well as porous wall.

Notation

$r(x)$ = Radius of the cross-section at right angles to the axis of the cone.

$$\bar{r} = \frac{r(x)}{a}$$

$$Z = \frac{\theta^2}{\nu} \cdot \frac{U}{r} \cdot \frac{dr}{dx}$$

a = representative length.

Momentum Integral Equation

Let x be measured along the meridian section of the surface and y along the outward drawn normal. Let u, v denote the velocities in the direction of x, y and $r(x)$ the distance measured from the axis to the surface of the body.

The boundary layer equation in the x-direction for three dimensional steady laminar incompressible flow is

$$u\frac{\partial u}{\partial x} + v\frac{\partial u}{\partial y} = -\frac{1}{\rho}\frac{\partial p}{\partial x} + \nu\frac{\partial^2 u}{\partial y^2} \qquad (1)$$

and the equation of continuity is

$$\frac{\partial (ur)}{\partial x} + v\frac{\partial (vr)}{\partial y} = 0$$

with continuous suction at the surface the boundary conditions are

$$\left.\begin{aligned} y = 0 &: \ u = 0,\ v = Vs \\ y = \infty &: \ u = U(x) \end{aligned}\right\} \qquad (2)$$

where V_s is the normal velocity at the surface and $U(x)$ is the potential flow velocity.

Assuming that the external pressure is impressed upon the boundary layer, the equation may be written as

$$u\frac{\partial u}{\partial x} + v\frac{\partial u}{\partial y} = u\frac{\partial u}{\partial x} + \nu\frac{\partial^2 u}{\partial y^2} \qquad (3)$$

Integrating with respect to y from $y = 0$ to $y = h$, where the layer $y = h$ is everywhere outside the boundary layer, we obtain.

$$\int_0^h \left[u\frac{\partial u}{\partial x} + v\frac{\partial u}{\partial y} - u\frac{\partial u}{\partial x} \right] = \nu\int_0^h \frac{\partial^2 u}{\partial y^2}\,dy \qquad (4)$$

from the equation (1)

$$\int_0^y \partial(vr) = \int_0^y \frac{\partial (ur)}{\partial x}\,dy$$

i.e. $$V\ r - V_s\ r = -\int_0^y \frac{\partial (ur)}{\partial x}\,dy$$

or, $$V = -\frac{1}{r}\int_0^y \frac{\partial (ur)}{\partial x}\,dy + Vs$$

Substituting for v into equation (4), we get

$$\int_0^h \left[u\frac{\partial u}{\partial x} + \frac{\partial u}{\partial y}\left\{ -\frac{1}{r}\int_0^y \frac{\partial (ur)}{\partial x}dy + V_s \right\} - U\frac{\partial u}{\partial x} \right] dy$$

$$= \nu\left[\frac{\partial u}{\partial x}\right]_0^h = -\frac{\tau_0}{\rho} \qquad (5)$$

Since $\dfrac{\partial u}{\partial y} = 0$, when $y = h$

and $\tau_0 = \mu\left[\dfrac{\partial u}{\partial y}\right]_y = 0$

The momentum integral equation

$$\frac{dt^*}{d\bar{x}} = \frac{2}{\bar{U}}\left[\ell - (2 + H)\wedge + \lambda - Z\right] \qquad (6)$$

When $t^* = \left(\dfrac{\theta}{a}\right)^2 \cdot \dfrac{U_0 a}{\nu}$ t*

Kinetic Energy Integral Equation

Adding $\dfrac{u}{2r}\left[\dfrac{\partial (ur)}{\partial x} + \dfrac{\partial (vr)}{\partial y}\right]$ to the left hand side of the equation (3) and multiplying through by u we get

$$u^2\frac{\partial u}{\partial x} + uv\frac{\partial u}{\partial y} + \frac{u^2}{2r}\left[\frac{\partial (ur)}{\partial x} + \frac{\partial (vr)}{\partial y}\right]$$

$$= uv\frac{\partial u}{\partial x} + \nu u\frac{\partial^2 u}{\partial y^2}$$

or, $\frac{3}{2}u^2\frac{\partial u}{\partial x}+\frac{u^2}{2}\frac{\partial v}{\partial u}+uv\frac{\partial u}{\partial y}+\frac{u^3}{2r}\frac{\partial r}{\partial x}$

$$=uv\frac{\partial u}{\partial x}+\nu u\frac{\partial^2 u}{\partial y^2}$$

or, $\frac{3}{2}u^2\frac{\partial u}{\partial x}+\frac{1}{2}\frac{\partial}{\partial y}(u^2 v)+\frac{u^3}{2r}\frac{\partial r}{\partial x}$

$$=uv\frac{\partial u}{\partial x}+\nu u\frac{\partial^2 u}{\partial y^2}$$

Integrating with respect to y from $y = 0$ to $y = h$

$$\int_0^h \frac{3}{2}u^2\frac{\partial u}{\partial x}dy+\frac{1}{2}\left[u^2 v\right]_0^h+\int_0^h\frac{u^3}{2r}\frac{dr}{dx}dy-\int_0^h\frac{u}{2}\frac{du^2}{dx}dy$$

$$=\left[\nu u\frac{\partial u}{\partial x}\right]_0^h-\int_0^h \nu\left(\frac{\partial u}{\partial x}\right)^2 dy \qquad (7)$$

Introducing further

$$H_\in=\frac{\in}{\theta}$$

and $\quad Z=\frac{\theta^2}{\nu}\cdot\frac{\text{U}}{r}\cdot\frac{dr}{dx}$

We get,

$$\frac{d}{dx}\left[\frac{\in^2}{\nu}\right]=\frac{2}{\text{U}}\left[2DH_\in-H_\in^2 Z+H_\in\lambda-3H_\in^2\wedge\right]$$

$$=\frac{2H_\in^2}{\text{U}}\left[2D-H_\in\left\{3\wedge+z\right\}\lambda\right] \quad (8)$$

The equation of K.E. Integral equation is

$$\frac{dH_\in}{d\bar{x}}=\frac{1}{\bar{\text{U}}t^*}\left[2D-H_\in\left\{\ell-(H-1)\wedge+\lambda\right\}+\lambda\right] (9)$$

Compatibility Condition at the Wall

At the surface of the body where $u = 0$ and $v = vs$, the boundary layer equation

$$u\frac{\partial u}{\partial x}+v\frac{\partial u}{\partial y}=U\frac{\partial u}{\partial x}+\nu\frac{\partial^2 u}{\partial y^2}$$

for the velocity profiles of H. Schlichting takes the form

$$K=\frac{\frac{\delta^2}{\theta^2}\wedge-\frac{\delta}{\theta}\lambda-1}{1+\frac{\delta}{\theta}\lambda(1-\frac{\pi}{6})} \qquad (10)$$

Conically Source Flow

It is proposed to investigate the effect of uniform suction on the boundary layer over the conically diverging wall. The conical wall beging from $\bar{x}=1$ where $\bar{x}=x/a$ and a is a certain representative length.

Sina = $r(x)/x$

where a is semi-vertical angle of the cone and $r(x)$ is radius of the circular section at a distance x from o at right angles to the axis of the cone.

or $r(x) = x$ W a = xC

When C = W, when $x = a$, $r(x) = ac$

The velocity at the leading edge is U_0

Therefore,

$U(x)$ = $U0a2c2$ / $x2c2$=$U0$a2/x2= $U0$ / $\bar{x}^2$

or, $\quad \bar{U}=\frac{U}{U_0}=\frac{1}{\bar{x}^2}$

$$\wedge=\frac{\theta^2}{\nu}\frac{du}{dx}=t^*\frac{d\bar{u}}{d\bar{x}}=\frac{-2t^*}{\bar{x}^3}$$

$$t^*=\left(\frac{\theta}{a}\right)^2\frac{U_0 a}{\nu}$$

$$\lambda=\frac{V_s\theta}{\nu}=\bar{V}_s t^{*\frac{1}{2}}$$

The momentum integral equation (6), Kinetic energy integral equation (9) and the wall comparability condition equation (10) for the conically source flow along porous wall.

$$\frac{dt^*}{d\bar{x}}=f(\bar{x},t^*,H)=2\bar{x}^2$$

$$\left[\ell + \frac{2t^*}{\bar{x}^3}(H+2) - \frac{t^*}{\bar{x}^3} + \bar{V}_s t^{*\frac{1}{2}}\right] \qquad (11)$$

$$\frac{dH_\in}{d\bar{x}} = \frac{\bar{x}^2}{t^*}$$

$$\left[2D - H_\in \left\{\ell + \frac{2t^*}{\bar{x}^3}(H-1) + \bar{V}_j t^{*\frac{1}{2}}\right\} + \bar{V}_j t^{*\frac{1}{2}}\right] (12)$$

$$(K+1)\left(\theta_j\right)^2 \left[1 + (1 - \frac{\pi}{6})K\right] \frac{\theta}{\delta} \bar{V}_s t^{*\frac{1}{2}} + \frac{2t^*}{\bar{x}^3} = 0 \quad (13)$$

The leading edge ($x = a$) is taken as the initial point for calculation. At the starting point $t^* = 0$ and $\lambda = \bar{V}_S t^{*\frac{1}{2}}$ vanishes.

Therefore at the initial point the values are

$$\bar{x} = 1$$

$$\bar{U} = 1$$

$$t^* = 0, \ \wedge = 0, \ \lambda = 0$$

$$k = -1$$

Corresponding to $k = -1$

$$\frac{\theta}{\delta} = 0.410$$

$$H = 2.660$$

$$I = 0.215$$

$$H_\in = 1.553$$

$$D = 0.169$$

Discussion of the Results

The momentum integral equation (11), the kinetic energy integral equation (12) and the wall compatibility condition (13) have been utilized for the numerical computation of these equations with the aid of Schlichting's profile for obtaining the point of separation for boundary layer along a permeable wall with uniform suction in a conically source flow of incompressible fluid.

Calculation have been made by employing Runga-Kutta method and Adom's quadrature formula.

The point of separation for suction velocity parameter $\bar{V}s = -.2$ and .4 are obtained at $\bar{x}(orx_S) = 1.1123$ and 1.1244 respectively.

The point of separation moves further drawn stream for increasing rate of suction parameter $\bar{V}s$.

As a special case, the problem reduces to the conically source flow along a solid wall by putting $\bar{V}s = 0$.

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Methods of Teaching and Learning of Secondary Mathematics

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Abstract

What is the best method of teaching a certain topic of Mathematics? How teacher can enable students to learn Mathematics? These are some of the basic questions for which every teacher want to find a solution. Though teaching is an art, no teacher can do a thoroughly good job of teaching Mathematics without a careful analysis, guide and method. The traditional methods of teaching are no longer adequate to meet the demands of modern Mathematics education. The issue of appropriate teaching methodology in Mathematics classroom is an evergreen topic which cannot be wished way. It is therefore important to search for simple methods of teaching and learning by which teachers could continuously inspire positive attitude of students in learning Mathematics. In this article procedure of teaching mathematics has been discussed.

Key Words: - Teaching and Learning.

Introduction

One of the most important properties of student is that they have a learning ability through the effect of environment and inherited intelligence. Civilized Society continuously trying to improve the teaching and learning methods in Mathematical education systems. Various studies were conducted to determine the most suitable teaching and learning style for the teachers as well as students.

Recently in various countries, secondary Mathematics classrooms have been dramatically changed in instructional approaches. Many teachers are focusing more on student oriented, exploration oriented activities in their classroom. It also seen that many Mathematics teachers attempted to improve their teaching methods with a belief that student centered instruction is the best way to achieve their goals of attaining students understanding of Mathematical concepts.

According to national education policy (1986), Mathematics should be visualized with the introduction of computer in school and teaching of Mathematics should be suitably redesigned with modern technological devices so that learning take place effectively.

Literature Review

Lave and wenger (1991), [5] viewed, becoming a full participant in the community of secondary Mathematics teaching involves engaging with the everyday discourse of practicing teachers and actively building relationships in that community by doing things together with Mathematics teachers.

P.C. Duchastel and P.F Merril (1973), [4] opined that teaching objectives would certainly make no difference if students cannot be made alert with their sensitivity to the learning situation.

M.K. Akinsola (1994),[2] stated that instructional method of teaching employed in the Mathematics classroom play a central role in developing students positive attitude towards Mathematics learning.

Methods of Teaching

There are many different ways in which effective teaching can perform. Some of the methods of teaching Mathematics have been discussed below:-

(i) Lecture Method:

The lecture method is the most widely used form of presentation. Lectures are used to introduce new topics, summarizing ideas, showing relationship between theory and practical, re-emphasizing main points. Finally, lectures can be effectively combined with other teaching methods to give added meaning and direction. Through this method a fruitful teaching of Mathematics can be expected if the teacher obtained the preparation in Planning, Rehearsing, Suitable language, Tone and face, Use of note etc.

(ii) Inductive-Deductive Method:

Inductive Method: Induction is that form of reasoning in which a general law is derived from a study of particular object or specific process. Student uses measurements, manipulators or constructive activities and pattern etc to discover a relationship.

Deductive Method: It is opposite of inductive method. Here the learner proceeds from general to particular, abstract to concrete, formula to example.

(iii) Project Method

This method aims to being practically designed experience in to the classroom. Often conducted over a period of three to six months the project gives an opportunity to work in a team environment and apply theory learned in the classroom.

(iv) Heuristic Method

Here the child is put in the place of discover. It involves finding out by the student by complete self-activity. The teacher is only passive observer. This method may be applied when the member of students are very less as it requires individual attention to each child.

(v) Analytic-Synthetic Method

Analytic method: It proceeds from unknown to known. "Analysis" means breaking up of the problem in hand so that it ultimately gets connected with something obvious or already known.

Synthetic Method: It is opposite of analytic method. Here one proceeds from known to unknown. It starts with something already known and connects that with the unknown parts of the statements.

(vi) Laboratory Method

It is more elaborate and practical form of the inductive Method. It makes the subject Mathematics more interesting as it combines play and activity.

The construction work in geometry is on the whole a laboratory work e.g. the drawing of a line, construction of an angle, construction of a triangle etc.

(vii) Dogmatic Method

In this method the rules and formulae are given to the class to cram. The teacher tells the pupils what to do, what to observe, how to attempt and how to conclude. Teacher works out the model sums on blackboard and students have merely to follow the patterns.

Technique of Teaching

For teaching Mathematics, one or more of the above stated methods can be applied. But learning of Mathematics mostly depends on how these methods are used or in which technique these methods are applied. Here stressed has been given on applying appropriate technique to use the methods of teaching Mathematics. Mathematics can be made more enjoyable and interesting subject. Here the chapter Quadratic Equation has been consider and through its application the way of teaching Mathematics has been explained so that the student feel easy and start loving the subject and enjoy it and also they become curious to know what is going to be happen.

Topic: Quadratic Equation

In this lesson we will discuss about Quadratic Equation and different methods of solving them.

Objective of the Lesson

At the end of this lesson, we will be able to :

1) Define a Quadratic Equation
2) Solve a Quadratic Equation using one of the following methods:
   i) Factorization
   ii) Converting into perfect square
   iii) Quadratic formula
   iv) Graphical representation

Introduction

Let us discuss the lesson with a problem.

Suppose the present age difference of John and Henry is 6 years. 10 years later the product of their ages will be 280. How old are John and Henry now?

Let us assume, John's present age = x years

So, Henry's present age = (6+x) years

According to the problem,

10 years later:

John's age will be = (x + 10) years

And Henry's age will be = (x + 6) + 10 years = (x+ 16) years

And product of their ages will be 280,

i.e. $(x + 10)(16 + x) = 280$

On solving this equation we get,

$x^2 + 26x - 120 = 0$

This is the equation with degree of variable 2.

Definition: A Quadratic Equation is a polynomial where the highest degree of a variable is 2.

The general form of Quadratic Equation in one variable is

$ax^2 + bx + c = 0$, where, a, b and c are Real numbers with $a \neq 0$

Note: The values that satisfy a quadratic equation are known as its zero, roots or solution of that quadratic equation.

Methods of solving Quadratic Equation

There are four methods of solving quadratic equation :

- Factorization
- Completing the square
- Quadratic Formula
- Graphical Representation

We will discuss now each of these methods in detail through the scenario about John and Henry's present ages.

Let us first use ***Factorization Method***. Consider the Quadratic Equation

$$x^2 + 26x - 120 = 0$$

To solve this equation we need to find the value of x.

We can solve this equation in several steps.

Step 1: Multiply the first and last term of the equation.

The product of the first and last term will be

$$-120x^2$$

Step 2 : we need to find the factors of $-120x^2$ in such a way that when we add them the sum of the difference is equal to the middle term $26x$

Step 3: we have the equation as :

$$x^2 + 30x - 4x - 120 = 0$$

Now we take a common factor x from first two terms and - 4 from last two terms,

$$x(x + 30) - 4(x + 30) = 0$$

Step 4 : Take $(x + 30)$ as a common factor from the equation.

Therefore the equation become

$$(x + 30)(x - 4) = 0$$

Therefore, $(x + 30) = 0$ or $(x - 4) = 0$

Thus the roots of the quadratic equation become ,

$$x = -30 \quad \text{or} \quad x = 4$$

As a person's age cannot be negative, the solution to the problem is

$$x = 4$$

Therefore, John's present age = $x = 4$ years

And Henry's present age = $6 + x = 6 + 4 =$ 10 years

Now let us discuss about the method ***Converting into Perfect Square***. Let us verify the method by considering the quadratic equation

$$x^2 + 26x \text{ “ } 120 = 0$$

Step 1: Bring the constant value 120 into right hand side of the equation

i.e. $x^2 + 26x = 120$

Step 2 : Divide the 2nd term $26x$ into two equal parts, (here each part is equal to $13x$)

i.e. $x^2 + 13x + 13x = 120$

Step 3: To make the equation a perfect square add $(13)^2$ to both sides of the equation, we get

$$x^2 + 13x + 13x + (13)^2 = 120 + (13)^2$$

$$\Rightarrow x^2 + 2 \times x \times 13 + (13)^2 = 120 + (13)^2$$

$$\Rightarrow (x + 13)^2 = 289$$

$$\Rightarrow (x + 13) = \pm 17$$

Solving the equation, we get

$(x + 13) =$ “ 17 or $(x + 13) = 17$

Hence the values of $x =$ “ 30 and 4

Therefore, answer to the problem is

John's age = $x = 4$ years

Henry's age = $(6 + x) = 6 + 4 = 10$ years

Sometime we may be found difficult to solve quadratic equation in converting the equation into perfect square. In that situation we can use Quadratic Formula to solve the equation.

Now let us discuss about ***Quadratic Formula***. To discuss about the Quadratic Formula let us consider the general form of a quadratic equation,

$$ax^2 + bx + c = 0$$

Step 1: Bringing the constant C to the right hand side of the equation, we get

$$ax^2 + bx = -c$$

Step 2: Dividing both sides of the equation with the co-efficient of x^2 i.e. a, we get,

$$x^2 + \frac{bx}{a} - \frac{c}{a}$$

Step 3: Multiplying and dividing the last term by 2 to get the equation in terms of 2ab, we get

$$x^2 + 2 \times x \times \frac{b}{2a} = \frac{-c}{a}$$

Step 4: We need to convert the equation into the form of identity , i.e.

$$(a + b)^2 = a^2 + 2ab + b^2$$

To do this we need to add $(b/2a)^2$ to both sides of the equation, we have

$$x^2+2\times x \ \times\frac{b}{2a}+(b/2a)^2=\frac{-c}{a}+(b/2a)^2$$

$- 4ac + B^2$

$\Rightarrow (x + b/2a)^2 \quad = 4a^2$

On solving the equation, we get

$$\Rightarrow \quad x = \frac{-b \pm \sqrt{b^2-4ac}}{2a}$$

Therefore, the roots of the equation are

$$x = \frac{-b-\sqrt{b^2-4ac}}{2a} \text{ and}$$

$$x = \frac{-b+\sqrt{b^2-4ac}}{2a}$$

We can also apply this quadratic formula to find the present ages of John and Henry.

Advantage of Quadratic Formula

Without solving the equation, we can calculate

$\Rightarrow$ The sum of the roots

$\Rightarrow$ The product of the roots

Application

a) Sum of the Roots :

Let the roots of the quadratic equation : $ax^2+bx+c=0$ be $\boldsymbol{\alpha}$ and $\boldsymbol{\beta}$ respectively.

$$a=\frac{-b-\sqrt{b^2-4ac}}{2a} \text{ and } \beta=\frac{-b-\sqrt{b^2-4ac}}{2a}$$

The sum of the roots of the quadratic equation is

$$\alpha + \beta \ = \ a+\beta=\frac{-b-\sqrt{b^2-4ac}}{2a}+$$

$$\frac{-b+\sqrt{b^2-4ac}}{2a}=\frac{-b}{a}$$

Therefore, the sum of the roots =

$$\frac{\text{- x coefficient}}{\text{x}^2\text{ coefficient}}$$

b) Product of the Roots:

$$a\times\beta= \quad \frac{-b-\sqrt{b^2-4ac}}{2a}\times\frac{-b+\sqrt{b^2-4ac}}{2a}=\frac{c}{a}$$

(on simplification)

i.e. Product of roots $=\dfrac{cons\tan t\ term}{x^2\ coefficient}$

Discriminant:

The roots obtained, depends on (b^2-4ac), which is known as the discriminant of a quadratic equation and it is denoted by

$$\text{D or } \triangle \ = b^2-4ac$$

We can use discriminant to find out the nature of the roots.

Nature of the Roots

- If $D > 0$ then the roots are real and distinct
- If $D > 0$ then the roots are imaginary
- If $D > 0$ then the roots are real and equal

As mentioned earlier we can also solve quadratic equation graphically.

Let us consider, ***Graphical Representation Method***. To find John and Henry's present age, consider the quadratic equation

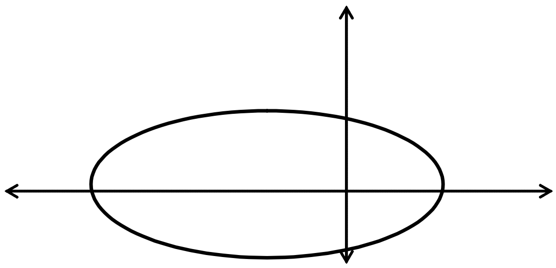
$$x^2 + 26x \text{ “ } 120 = 0$$

Let us assume, $Y = x^2 + 26x$ “ 120

We can see the table displayed on ordered pair below based on the equation: Y = x2 + 26x – 120

X	−40	−30	−20	−10	0	10
Y	440	0	−240	−280	−120	240

On ploting the order pairs on the graph, the curve of the quadratic equation: $y=x^2+26x-120$, images as a parabola.



This parabola touches the X-axis at (– 30,0) and (4, 0).

Since this parabola touches the X-axis at – 30 and 4 , therefore , the roots of the equation

$x^2 + 26x - 120 = 0$ are – 30 and 4

Therefore, John's age = x = 4 years

Henry's age = 6 + x = 6 + 4 = 10 years.

Note : The curve of the parabola varies, based on

- ➢ The values of the coefficient of x^2, i.e. " a"
- ➢ The discriminant ($b^2 - 4ac$)

Impact:

In a quadratic polynomial $ax^2 + bx + c$

- ⇨ If a > 0 , then the parabola opens upward.
- ⇨ If a < 0 , then the parabola open downwards.

Impact of Discriminant (D = b^2 " 4ac) :

- ⇨ If D > 0 , then the parabola intersects the X –axis at two different points.
- ⇨ If D < 0 , then the parabola does not intersects the X –axis.
- ⇨ If D = 0 , then the parabola intersects the X –axis at one point only.

Summary of the Lesson:

- ✓ The roots of a quadratic equation of the form $ax^2 + bx + c = 0$, can be calculated using the formula:

$$x = \frac{-b \pm \sqrt{b^2 - 4ac}}{2a}$$

- ✓ The sum of the roots of a quadratic equation is : $\alpha + \beta = \frac{-b}{a}$
- ✓ The Product of the roots of a quadratic equation is : $\alpha\beta = \frac{c}{a}$
- ✓ In a quadratic equation ($b^2 - 4ac$) is called discriminant and it is denoted by D or Δ.
- ✓ The nature of the roots of a quadratic equation can be summed up as follows :
  - If D > 0 then the roots are real and distinct.
  - If D > 0 then the roots are imaginary.
  - If D > 0 then the roots are real and equal.

Conclusion

Different methods of teaching Mathematics and its implementation have been proposed by different experts and the knowledge of these methods may help in working out a better teaching strategy. It is not appropriate for a Mathematics teacher to commit to one particular method. A teacher should adopt a teaching approach after considering the nature of the students, their interest and maturity, classroom environment and the resources available for teaching Mathematics. Every method has certain merits and few demerits and it is the work of the mathematics teacher to decide which method is best for the students to provide a worthwhile and effective teaching.

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Use of Congruence in the Introduction of Cryptography

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Abstract

As we know that cryptography is the only known practical means for protecting information through public communications networks, such as those using telephone lines, microwave or satellites In this topic we shall how congruence is used in cryptography.

Introduction

Cryptography is a Greek kryptos meaning hidden and graphein meaning to write thus cryptography is a science of making communications unintelligible to all except authorized parties. In the language of cryptography, where codes are called ciphers, the information to be concealed is called plaintext. After transformation to a secret from, a message is called ciphertext. The process of converting from plaintext to ciphertext is said to be encrypting (or enciphering,) whereas the reverse process of changing from ciphertext back to plaintext is called decrypting (or deciphering)

One of the earliest cryptographic systems was used by the great Roman emperor Julius Caesar around 50 B.C. Caesar wrote to Marcus Cicero using a rudimentary substitution cipher in which each letter of the alphabet is replaced by the letter that occurs three places down the alphabet, with the last three letters cycled back to the first three letters. If we write the ciphertext equivalent underneath the plaintext letter, the substitution alphabet for the Caesar cipher is given by

Plaintext:

ABCDEFGHIJKLMNOPQRSTUVWXYZ

Ciphertext:

DEFGHIJKLMNOPQRSTUVWXYZABC

For example, the plaintext message

CAESAR WAS GREAT

Is transformed into the ciphertext

FDHVDU ZDV JUHDW

The Caesar cipher can be described easily using congruence theory. Any plaintext is first expressed numerically by translating the characters of the text into digits by means of some correspondence such as the following:

A	B	C	D	E	F	G	H	I	J	K	L	M
00	01	02	03	04	05	06	07	08	09	10	11	12
N	O	P	Q	R	S	T	U	V	W	X	Y	Z
13	14	15	16	17	18	19	20	21	22	23	24	25

If P is the digital equivalent of a plantext letter and C is the digital equivalent of the corresponding ciphertext letter, then

$C \equiv + P + 3 \pmod{26}$

Thus, for instance, the letters of the message in Eq. (1) are converted to their equivalent:

02 00 04 18 00 17 22

00 18 06 17 04 00 19

Using the congruence $C \equiv + P + 3 \pmod{26}$, this becomes the ciphertext

05 03 07 21 03 20 25 03

21 09 20 07 03 22

To recover the plaintext, the procedure is simply reversed by means of the congruence $P \equiv + C - 3 \quad C + 23 \pmod{26}$

The Caesar cipher is very simple and, hence, extremely insecure. Caesar himself soon abandoned this scheme – not only because of its insecurity, but also because he did not trust Cicero, with whom he necessarily shared the secret of the cipher. An encryption scheme in which each letter of the original massage is replaced by the same cipher substitute is known as a monoalphabetic cipher. Such cryptographic systems are extremely vulnerable to statistical methods of attack because they preserve the frequency, or relative commonness, of individual letters. In a polyalphabetic cipher, a plaintext letter has more than one ciphertext equivalent: the letter E, for instance, might be represented by J, Q, or X, depending on where it occurs in the message.

General fascination with cryptography had its initial impetus with the short story the fold Bug, published in 1843 by the American writer Edgar Allan Poe.

It is a fictional tale of the use of a table of letter frequencies to decipher directions for finding captain Kidd's buried treasure. Poe fancied himself a cryptologist far beyond the ordinary. Writing for Alexander's Weekly, a Philadelphia newspaper, he once issued claim that he could solve "forthwith" any monoalphabetic substitution cipher sent in by readers. The Challenge was taken up by one G. W. Kulp, who submitted a 43 word ciphertext in longhand. Poe showed in a subsequent column that the entry was not genuine, but rather a "jargon of random characters having no meaning whatsoever." When Kulp's cipher submission was finally decoded in 1975, the reason for the difficulty became clear; the submission contained a major error on kulp's part, along with 15 minor errors, which were most likely printer's mistake in reading Kulp's longhand.

The most famous example of a polyalphabetic cipher was published by the French cryptographer Blaise de Vigenere (1523-1596) in his Traicte de Chiffres of 1586. To Implement this system, the communicating parties agree on an easily remembered word or phrase. With the standard alphabet numbered from A= 00 to Z = 25, the digital equivalent of the keyword is repeated as many times as necessary beneath that of the plaintext message. The message is then enciphered by adding, modulo 26, each plaintext number to the immediately beneath it. The process may be illustrated with the keyword READY, whose numerical version is 17 04 00 03 24. Repetitions of this sequence are arranged below the numerical plaintext of the message.

ATTACK AT ONCE

To produce the array

00	19	19	00	02	10	00	19	14	13	02	04
17	04	00	03	24	17	04	00	03	24	17	04

When the columns are added modulo 26, the plaintext message is encrypted as

17	23	19	03	00	01	04	19	17	11	19	08

Or, converted to letters,

RXTDAB ET RLTI

Notice that a given letter of plaintext is represented by different letters in ciphertext. The double t in the word ATTACK no longer appears as a double letter when ciphered, while the ciphertext letter R first corresponds to A and then to O in the original message.

In general, any sequence of n letters with numerical equivalents $b_1, b_2, \ldots\ldots B_n$ ($00 \leq b_i \leq 25$) will serve as the keyword. The plaintext message is expressed as successive blocks $p_1 p_2 \ldots p_n$ of n two-digit integers p_i and then converted to ciphertext blocks $c_1 c_2 \ldots C_n$ means of the congruences

$$C_i \equiv P_i + b_i \pmod{26}$$

Decryption is carried out by using the relations

$$P \equiv C_i - b_i ((\mathrm{mod}26)\ 1 \leq i \leq n$$

A weakness in Vigener's approach is that once the length of the keyword has been determined, a coded message can be regarded as a number of separate monoalphabetic ciphers, each subject to straightforward frequency analysis. A variant to the continued repetition of the keyword is what is called a running key, a random assignment of ciphertext letters to plaintext letters to plaintext letters. A favorite procedure for generating such keys is to use the text of a book, where both sender and receipient know the title of the book and starting point of the appropriate lines. Because a running key cipher completely obscures the underlying structure of the original message, the system was long thought to be secure. But it does not, as Scientific American once claimed, produce ciphertext that is "impossible of translation,"

A clever modification that vigenere contrived for his polyaphabetic cipher is currently called the autokey ("automatic key"). This approach makes use of the plaintext message itself in constructing the encryption key. The idea is to start off the keyword with a short seed of primer (generally a single letter) followed by the plaintext, whose ending is truncated by the length of the seed. The autokey cipher enjoyed considerable popularity it the 16th and 17th centuries, since all it required of legitimate pair of users was to remember the seed, which could easily be changed.

Let us given a simple example of the method.

Example - Assume that the message

ONE ID BY DAWN

Is to be encrypted. Taking the letter K as the seed, the keyword becomes

KONEIFBYDAW

When both the plaintext and keyword are converted to numerical form, we obtain the array

14	13	04	08	05	01	24	03	00	22	13
10	14	13	04	08	05	01	24	03	00	22

Adding the integers in matching positions modulo 26 yield the ciphertext

24 01 17 12 13 06 25 01 03 22 09

Or, changing back to letters:

YBR MN GZ BDWJ

Decipherment is achieved by returning to the numerical form of both the plaintext and its ciphertext. Suppose that the plaintext has digital equivalents $p_1\ p_2 \ldots . p_n$ and the ciphertext $c_1 c_2 \ldots$ c. If S indicates the seed, then the first plaintext number is

$$P_1 = c_1 - s = 24\text{-}10\ 14 \pmod{26}$$

Thus, the deciphering transformation becomes

$$P_k \equiv C_k - P_{k-1} \ (\text{mod}26),\ 2 \leq k \leq n$$

This recovers, for example, the integers

$$P_2 \equiv 01\text{-}\ 14 = -13 \equiv 13 \ (\text{mod}26)$$

$$P_3 \equiv 17\text{-}13 \equiv 4 \ (\text{Mod } 26)$$

Where, to maintain the two-digit format, the 4 is written 04.

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A Theoretical Study and Solution Chemistry of Transition Metal Complex of furfural.

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Abstract

The interaction between furfural and biologically important metal ions Co^{II} , Ni^{II} , Cu^{II} and Zn^{II} was studied in solutions phase through potentiometry and gas phase using ah initio HF/6-31G* and MP2/6-31G* methods.Ground state geometeries were optimized and tested through frequency analysis. For simplicity 1:1 species ratio was considered for theoretical study. It was found out that furfural is bidentate ligand and binds through furan oxygen and carbonyl oxygen with entire four metal ions in the gas and solution phases.The binding effect is found to be in the order - furan oxygen < carbonyl oxygen. The result indicate that furan oxygen binds to the metal ion through S and Pz orbitals , and the carbonyl oxygen binds through S and Py orbitals.The binding effects and modes of both the oxygen are found to differ. Furfural forms high spin complex with Co^{II} and Ni^{II} with multiplicity of 4 and 3 respectively. The strength of Cu^{II} – furfural binding in both gas and solution phase is found to be lesser than expected. This may be due to the weak binding effect of furan oxygen. The results demonstrate that the order of binding in solution phase is $Co^{II} < Ni^{II} < Cu^{II} > Zn^{II}$.

Keywords – Theoretical Study, Solution Study, Metal Complexes, Furfural, Schiff's Base.

Introduction

Furfural is a well known natural compound used extensively in food, fuel and paint industries[1-3].

Schiff's base complexes of Co^{II} , Ni^{II} , and Cu^{II} involving furfural and 2-aminopyridine, ethylenediamine, diethylenetriamine, dipropylenetriamine, spermidine , hydrazine , chloroaniline or 2-aminopyridine have been found to have antibacterial and anti fungal activities[4-9].

Schiff's base of furfural and anthranilic acid can act as antitumor agent[10].The present study reveals the interaction between furfural and the 3d-metals ions such as Co^{II}, Ni^{II}, Cu^{II} and Zn^{II}.

Experimental

The General Atoms and Molecular Electronic Structure System (GAMESS) program[11-12] was employed for the ab initio (HF) calculations made in this study. The 6-31G* basis set[1] – was used. Binary stability constants were determined by potentiometic method at 27 ± 1^0C and 0.1μ ionic strength. Initial geometry of the molecule was constructed from the standard geometry of furfural. The conformational search was carried out along the axis at 180^0. All the semiempirical calculations were carried out by MOPAC 6 program[14-15].

Result and Discussion

In order to have a clear understanding about the metal complexes of furfural, it is necessary to optimize its ground state equilibrium geometry. A systematic conformational study can help to provide a clear understanding of the energy and often important insights into equilibrium geometry. The conformers and the energy profile of furfural along the a-axis.

Considering the different structures for the molecule, there might be several possible structures arising from rotation around the C_4-C_6 bond.

The calculated relative conformational energy values, however, are often sensitive to method used in the calculations. In order to obtain the lowest energy con-former for the furfural the calculations were done at the RHF/6-31G* and MP2/6-31G*. On the basis of these methods it is able to distinguish among the conformers with significantly lower energy **trans** furfural. The relative energy difference of the studied conformers **cis** and **trans** furfural is in the order of 2 kcal/mol.

The energy profile is almost symmetrical. The perpendicular conformer is in the top of the barrier.The planar conformers are stable than non-planar conformers. For this system it was observed that the calculated barrier height is very large for the free rotation of aldehyde group, suggesting that furfural is a rigid molecule. The rotational barrier is due to loss of symmetry, loss of electron delocalization and non-bonded interaction between aromatic electron cloud of furan ring and carbonyl oxygen in furfural.

Relative distribution among the population of **cis** and **trans** conformers is 1:1. The calculated dipole moment values of the conformers show that the cis conformer has the higher dipole moment than **trans**. It has been reported that , conformer with higher dipole moment exist in solid and solution phases[16].Thus it has been concluded that the trans conformer is the theoretically expected one and the **cis** form exists in solid and solution phases. Since our studies are made in solution, the **cis** conformer only is considered for the present study. Further, NMR studies have documented that furfural exists in **cis** conformation in liquid[17-18].

Important geometrical parameters are given in Table-1. A good correlation exists between the computed and experimental geometrical parameters.

It is emphasized that bond angles have a mean deviation of 0-3^0. From the experimental values the C-C bond and C-O single bond length are determined to be 1.432 A^0 and 1.309 A^0.The C=C and C=N have a double bond length of 1.346 A^0 and 1.309 A^0 respectively for to 2-furfuraldoxime[19]. Taking in to account the effect of conjugation, the calculated values of the cis-furfural molecule is in reasonable agreement with the above mentioned experiment data. So the calculated values of cis-furfural are very reasonable. The molecule is planar with no imaginary frequency and it has C_S symmetry.

Among the ab initio calculations, MP2 level has lesser charge density for atoms than HF. This is due to the higher electron correlation of MP2 over HF.The O_3 and O are the higher negative centers for protonation and complexation. Mulliken atomic orbital population analysis shows an increase in electron population of the P_7 and a decrease in the population of the P_X and P_Y orbitals on the O_3 due to aromatization of furfural. In the semiempirical methods of %-orbitals has more population and P_X orbital has more electron density than P_Y in O_3 but it is a reverse in the ab initio methods.

In O_7 atom the S and P_Y orbital have more population in AM1 and PM3 methods. RHF and MP2 have more population in the P_X orbital. The results reveal that the coordination sites O_3 and O_7 possess different characters.

Important Geometrical Parameters of M^{II}-cis –furfural

Table-1

Bond length(O)	Co^{II} cis-furfural	Ni^{II} cis- furfural	Cu^{II} cis-furfural	Zn^{II} cis- furfural
C_1-C_2	1.3367	1.3360	1.4170	1.3365
C_2-O_3	1.4030	1.4053	1.3296	1.4049
O_3-C_4	1.4372	1.4390	1.3706	1.4394
C_4-C_5	1.3535	1.3532	1.4262	1.3549
C_4-C_6	1.4072	1.4080	1.4756	1.4064
C_6-O_7	1.2671	1.2678	1.2153	1.2710
M-O_3	2.0594	2.0090	4.5682	2.0090

Table-2

Bond angle(O)	Co^{II} cis-furfural	Ni^{II} cis- furfural	Cu^{II} cis-furfural	Zn^{II} cis- furfural
C_2-O_3-C_4	110.0236	109.8865	110.2992	109.9094
O_3-C_4-C_5	105.8890	105.9692	108.0306	106.0050
O_3-C_4-C_6	108.1988	108.0795	108.2245	107.9795
C_4-C_6-O_7	112.8944	113.0361	119.2253	113.4593
M-O_3-C_4	110.1334	109.3111	76.5915	108.2349

Table-3

Dihedral angle(O)	Co^{II} cis-furfural	Ni^{II} cis- furfural	Cu^{II} cis-furfural	Zn^{II} cis- furfural
C_1-C_2-O_3-C_4	0.0000	0.0000	0.0000	0.0000
C_2-O_3-C_4-C_5	0.0000	0.0000	0.0000	0.0000
C_2-O_3-C_4-C_6	180.0000	180.0000	180.0000	180.0000
O_3-C_4-O_6-0_7	180.0000	180.0000	180.0000	180.0000

Because of the computational expense of ab initio calculations the number of atoms which can be included is limited and so that the use of ab initio methods is restricted to the complexation of single cis-furfural ligand only and not for water molecules.

It was observed that cis-furfural has two electronegative atoms, furfural (O_3) and carboxyl (O_7) oxygens. Therefore m-furfural either binds monodentatedly through O_3 or O_7 oxygen or in bidentate mode through both O_3 and O_7 oxygen with metals ions. In order to find out the mode of binding of M^{II} with cis-furfural. Its HF energy values are calculated (Table-2). It was found that the bidentate mode of binding is preferred. The binding energy is in the order- O_3 and O_7 > O_7 > O_3.

The less binding energy of mondentate O_3 than O_7 is due to the aromatization of the O_3 electrons.

Thus cis-furfural is a bidentate ligand and binds M^{II} through furfural (O_3) and carbonyl (O_7) oxygens. It is expected that the Co^{II} and Ni^{II} form complexes with different multiplicities. The relative energy values for multiplicities of Co^{II} / Ni^{II}-cis-furfural are given table-3.

The Co^{II} / Ni^{II}-cis-furfural complexes with multiplicity of four and three are more stable than multiplicity of two and one respectively. This implies that Co^{II} / Ni^{II}-cis-furfural complexes are high spin.

The bond angle M-O_3-C_4 for Co^{II}, Ni^{II}, Cu^{II} and Zn^{II} are 110.1334^0, 109.3111^0, 76.5915^0 and 108.2349^0 respectively. Thus it has been concluded that Co^{II}, Ni^{II} and Zn^{II} form tetrahedral geometry with one molecule of furfural. Cu^{II} forms square planar in 1:1 complex with furfural.

In solution all the four metal ions form MA and MA_2 types of complexes with furfural in addition to HA species. The K_{10} values are 1.81, 2.02, 2.69 and 1.74 log units respectively for Co^{II}, Ni^{II}, Cu^{II} and Zn^{II} metals ions-furfural complexes. These values are comparable with the stability constant values of complexes of furan semicarbazones, in which the ligands are bidentate and bind through furfural and carbonyl oxygens[20]. It has been concluded that furfural behaves as a bidentate ligand and binds through furfural oxygen and carbonyl group of aldehyde with all the four metals ions. Bryson and Dwyer[21] potentiometrically studied a number of metal complexes of β-furfuraldoximates and found that the ligands bind through furfural and hydroxyl oxygens. In the solid complexes of furancemicarbazones, the furfural oxygen also takes part in the coordination[20].

The strength of Cu^{II}-furfural binding in both gas and solution phases is found to be lesser than expected. This may be due to the weak binding effect of furan oxygen. The stability of binding in solution phase follows Irving-William[22] order- $Co^{II} < Ni^{II} < Cu^{II} > Zn^{II}$.

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Basic Concepts in Linear Differential Equation: A New Approach

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Linear differential equations are of the form

$$Ly = f$$

where the differential operator L is a linear operator, y is the unknown function (such as a function of time y(t)), and the right hand side f is a given function of the same nature as y (called the source term). For a function dependent on time we may write the equation more expressively as

$$Ly(t) = f(t)$$

and, even more precisely by bracketing

$$L[y(t)] = f(t)$$

The linear operator L may be considered to be of the form

$$L_n(y) \equiv \frac{d^n y}{dt^n} + A_1(t)\frac{d^{n-1}y}{dt^{n-1}} + \cdots + A_{n-1}(t)\frac{dy}{dt} + A_n(t)y$$

The linearity condition on L rules out operations such as taking the square of the derivative of y; but permits, for example, taking the second derivative of y. It is convenient to rewrite this equation in an operator form

$$L_n(y) \equiv \left[D^n + A_1(t)D^{n-1} + \cdots + A_{n-1}(t)D + A_n(t)\right]y$$

where D is the differential operator d/dt (i.e. $Dy = y'$, $D^2y = y''$,...), and the A_n are given functions.

Such an equation is said to have order n, the index of the highest derivative of y that is involved. A typical simple example is the linear differential equation used to model radioactive decay. Let N(t) denote the number of radioactive atoms in some sample of material at time t. Then for some constant $k > 0$, the number of radioactive atoms which decay can be modelled by

$$\frac{dN}{dt} = -kN$$

If y is assumed to be a function of only one variable, one speaks about an ordinary differential equation, else the derivatives and their coefficients must be understood as (contracted) vectors, matrices or tensors of higher rank, and we have a (linear) partial differential equation.

The case where $f = 0$ is called a homogeneous equation and its solutions are called complementary functions. It is particularly important to the solution of the general case, since any complementary function can be added to a solution of the inhomogeneous equation to give another solution (by a method traditionally called *particular integral and complementary function*). When the *Ai are* numbers, the equation is said to have *constant coefficients*.

Homogeneous Equations with Constant Coefficients

The first method of solving linear ordinary differential equations with constant coefficients is due to Euler, who realised that solutions have the form e^{zx}, for possibly-complex values of z. The exponential function is one of the few functions that keep its shape after differentiation. In order for the sum of multiple derivatives of a function to sum up to zero, the derivatives must cancel each other out and the only way for them to do so is for the derivatives to have the same form as the initial function. Thus, to solve

$$y^{(n)} + A_1 y^{(n-1)} + \cdots + A_n y = 0$$

we set $y = e^{zx}$, leading to

$$z^n e^{zx} + A_1 z^{n-1} e^{zx} + \cdots + A_n e^{zx} = 0.$$

Division by e^{zx} gives the nth-order polynomial

$$F(z) = z^n + A_1 z^{n-1} + \cdots + A_n = 0.$$

This algebraic equation $F(z) = 0$, is the characteristic equation considered later by Gaspard Mongeand Augustin-Louis Cauchy.

Formally, the terms

$$y^{(k)} \qquad (k = 1,2,\ldots,n).$$

of the original differential equation are replaced by z^k. Solving the polynomial gives n values of $z, z_1, \ldots, z_n$. Substitution of any of those values for z into e^{zx} gives a solution $e^{z_i x}$. Since homogeneous linear differential equations obey the superposition principle, any linear combination of these functions also satisfies the differential equation.

When these roots are all distinct, we have n distinct solutions to the differential equation. It can be shown that these are linearly independent, by applying the Vandermonde determinant, and together they form a basis of the space of all solutions of the differential equation.

Examples

$$y'''' - 2y''' + 2y'' - 2y' + y = 0$$

has the characteristic equation

$$z^4 - 2z^3 + 2z^2 - 2z + 1 = 0.$$

This has zeroes, i, $-i$, and 1 (multiplicity 2). The solution basis is then

$$e^{ix}, e^{-ix}, e^x, xe^x.$$

This corresponds to the real-valued solution basis

$$\cos x, \sin x, e^x, xe^x.$$

The preceding gave a solution for the case when all zeros are distinct, that is, each has multiplicity 1. For the general case, if z is a (possibly complex) zero (or root) of $F(z)$ having multiplicity m, then, for $k \in \{0,1,\ldots,m-1\}$, $y = x^k e^{zx}$ is a solution of the ODE. Applying this to all roots gives a collection of n distinct and linearly independent functions, where n is the degree of $F(z)$. As before, these functions make up a basis of the solution space. If the coefficients A_i of the differential equation are real, then real-valued solutions are generally preferable. Since non-real roots z *then* come in conjugate pairs, so do their corresponding basis functions $x^k e^{zx}$, and the desired result is obtained by replacing each pair with their real-valued linear combinations $\mathrm{Re}(y)$ and $\mathrm{Im}(y)$, where y is one of the pair. A case that involves complex roots can be solved with the aid of Euler's formula.

Examples

Given $y'' - 4y' + 5y = 0$. The characteristic equation is $z^2 - 4z + 5 = 0$ which has roots $2+i$ and $2-i$. Thus the solution basis $\{y_1, y_2\}$ is $\{e^{(2+i)x}, e^{(2-i)x}\}$. Now y is a solution if and only if $y = c_1 y_1 + c_2 y_2$ for $c_1, c_2 \in \mathbb{C}$.

Because the coefficients are real,

- we are likely not interested in the complex solutions
- our basis elements are mutual conjugates

The linear combinations

$$u_1 = \mathrm{Re}(y_1) = \frac{y_1 + y_2}{2} = e^{2x}\cos(x) \text{ and}$$

$$u_2 = \mathrm{Im}(y_1) = \frac{y_1 - y_2}{2i} = e^{2x}\sin(x)$$

will give us a real basis in $\{u_1, u_2\}$.

Simple Harmonic Oscillator

The second order differential equation

$D^2y = -k^2y$,

which represents a simple harmonic oscillator, can be restated as

$(D^2 + k^2)y = 0$.

The expression in parenthesis can be factored out, yielding

$(D + ik)(D - ik)y = 0$,

which has a pair of linearly independent solutions, one for

$(D - ik)y = 0$

and another for

$(D + ik)y = 0$.

The solutions are, respectively,

$y_0 = A_0e^{ikx}$

and

$y_1 = A_1e^{-ikx}$.

These solutions provide a basis for the two-dimensional "solution space" of the second order differential equation: meaning that linear combinations of these solutions will also be solutions. In particular, the following solutions can be constructed

$$y_{0'} = \frac{A_0e^{ikx} + A_1e^{-ikx}}{2} = C_0\cos(kx)$$

and

$$y_{1'} = \frac{A_0e^{ikx} - A_1e^{-ikx}}{2i} = C_1\sin(kx).$$

These last two trigonometric solutions are linearly independent, so they can serve as another basis for the solution space, yielding the following general solution:

$y_H = C_0\cos(kx) + C_1\sin(kx)$.

Damped Harmonic Oscillator

Given the equation for the damped harmonic oscillator:

$$\left(D^2 + \frac{b}{m}D + \omega_0^2\right)y = 0,$$

the expression in parentheses can be factored out: first obtain the characteristic equation by replacing D with λ. This equation must be satisfied for all y, thus:

$$\lambda^2 + \frac{b}{m}\lambda + \omega_0^2 = 0.$$

Solve using the quadratic formula:

$$\lambda = \frac{-b/m \pm \sqrt{b^2/m^2 - 4\omega_0^2}}{2}.$$

Use these data to factor out the original differential equation:

$$\left(D + \frac{b}{2m} - \sqrt{\frac{b^2}{4m^2} - \omega_0^2}\right)\left(D + \frac{b}{2m} + \sqrt{\frac{b^2}{4m^2} - \omega_0^2}\right)y = 0.$$

This implies a pair of solutions, one corresponding to

$$\left(D + \frac{b}{2m} - \sqrt{\frac{b^2}{4m^2} - \omega_0^2}\right)y = 0$$

and another to

$$\left(D + \frac{b}{2m} + \sqrt{\frac{b^2}{4m^2} - \omega_0^2}\right)y = 0$$

The solutions are, respectively,

$$y_0 = A_0e^{-\omega x + \sqrt{\omega^2 - \omega_0^2}x} = A_0e^{-\omega x}e^{\sqrt{\omega^2 - \omega_0^2}x}$$

and

$$y_1 = A_1e^{-\omega x - \sqrt{\omega^2 - \omega_0^2}x} = A_1e^{-\omega x}e^{-\sqrt{\omega^2 - \omega_0^2}x}$$

where $\omega = b / 2m$. From this linearly independent pair of solutions can be constructed another linearly independent pair which thus serve as a basis for the two-dimensional solution space:

$$y_H(A_0, A_1)(x) = \left(A_0 \sinh\sqrt{\omega^2 - \omega_0^2}\,x + A_1 \cosh\sqrt{\omega^2 - \omega_0^2}\,x\right)e^{-\omega x}.$$

However, if $|\omega| < |\omega_0|$ then it is preferable to get rid of the consequential imaginaries, expressing the general solution as

$$y_H(A_0, A_1)(x) = \left(A_0 \sin\sqrt{\omega_0^2 - \omega^2}\,x + A_1 \cos\sqrt{\omega_0^2 - \omega^2}\,x\right)e^{-\omega x}.$$

This latter solution corresponds to the underdamped case, whereas the former one corresponds to the overdamped case: the solutions for the underdamped case oscillate whereas the solutions for the overdamped case do not.

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The Impact of Mineral Toxicity Stress in Plant

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Soil pH or soil reaction is an indication of the acidity or alkalinity of soil and is measured in pH units. Soil pH is defined as the negative logarithm of the hydrogen ion concentration. The pH scale goes from 0 to 14 with pH 7 as the neutral point. As the amount of hydrogen ions in the soil increases the soil pH decreases thus becoming more acidic. From pH 7 to 0 the soil is increasingly more acidic and from pH 7 to 14 the soil is increasingly more alkaline or basic.

Descriptive terms commonly associated with certain ranges in soil pH are:

- extremely acid, < than 4.5; lemon=2.5; vinegar=3.0; stomach acid=2.0; soda=2–4
- very strongly acid, 4.5–5.0; beer=4.5–5.0; tomatoes=4.5
- strongly acid 5.1–5.5; carrots=5.0; asparagus=5.5; boric acid=5.2; cabbage=5.3
- moderately acid, 5.6–6.0; potatoes=5.6
- slightly acid, 6.1–6.5; salmon=6.2; cow's milk=6.5
- neutral, 6.6–7.3; saliva=6.6–7.3; blood=7.3; shrimp=7.0
- slightly alkaline, 7.4–7.8; eggs=7.6–7.8
- moderately alkaline, 7.9–8.4; sea water=8.2; sodium bicarbonate=8.4
- strongly alkaline, 8.5–9.0; borax=9.0
- very strongly alkaline, > than 9.1; milk of magnesia=10.5, ammonia=11.1; lime=12

Measuring Soil *pH*

Soil pH provides various clues about soil properties and is easily determined. The most accurate method of determining soil pH is by a pH metre. A second method which is simple and easy but less accurate then using a pH metre, consists of using certain indicators or dyes.

Phosphorus is never readily soluble in the soil but is most available in soil with a pH range centred around 6.5. Extremely and strongly acid soils (pH 4.0-5.0) can have high concentrations of soluble aluminum, iron and manganese which may be toxic to the growth of some plants. A pH range of approximately 6 to 7 promotes the most ready availability of plant nutrients.

But some plants, such as azaleas, rhododendrons, blueberries, white potatoes and conifer trees, tolerate strong acid soils and grow well. Also, some plants do well only in slightly acid to moderately alkaline soils. However, a slightly alkaline (pH 7.4-7.8) or higher pH soil can cause a problem with the availability of iron to pin oak and a few other trees in Central New York causing chlorosis of the leaves which will put the tree under stress leading to tree decline and eventual mortality.

The soil pH can also influence plant growth by its effect on activity of beneficial microorganisms Bacteria that decompose soil organic matter are hindered in strong acid soils. This prevents organic matter from breaking down, resulting in an accumulation of organic matter and the tie up of nutrients, particularly nitrogen, that are held in the organic matter.

Changes in Soil *pH*

Soils tend to become acidic as a result of: (1) rainwater leaching away basic ions (calcium, magnesium, potassium and sodium); (2) carbon dioxide from decomposing organic matter and root respiration dissolving in soil water to form a weak organic acid; (3) formation of strong organic and inorganic acids, such as nitric and sulfuric acid, from decaying organic matter and oxidation of ammonium and sulfur fertilizers. Strongly acid soils are usually the result of the action of these strong organic and inorganic acids.

Lime is usually added to acid soils to increase soil pH. The addition of lime not only replaces hydrogen ions and raises soil pH, thereby eliminating most major problems associated with acid soils but it also provides two nutrients, calcium and magnesium to the soil. Lime also makes phosphorus that is added to the soil more available for plant growth and increases the availability of nitrogen by hastening the decomposition of organic matter.

Liming materials are relatively inexpensive, comparatively mild to handle and leave no objectionable residues in the soil. Some common liming materials are: (1) Calcic limestone which is ground limestone; (2) Dolomitic limestone from ground limestone high in magnesium; and (3) Miscellaneous sources such as wood ashes. The amount of lime to apply to correct a soil acidity problem is affected by a number of factors, including soil pH, texture (amount of sand, silt and clay), structure, and amount of organic matter. In addition to soil variables the crops or plants to be grown influence the amount of lime needed.

Causes and Effects of Soil Acidity

Soil acidity is a crop production problem of increasing concern in central and western Oklahoma. Although acid soil conditions are more widespread in eastern Oklahoma, the more natural occurrence there has resulted in farm operators being better able to manage soil acidity in that part of the state.

However, in central and western Oklahoma the problem appears to grow with time. This fact sheet explains why soils become acid and the problems acid soils create for crop production. OSU Extension Facts No. 2229 explains how soil acidity and the lime requirement are determined by soil testing. A subsequent fact sheet discusses managing wheatland soils in Oklahoma.

Why Soils are Acid

The four major causes for soils to become acid arelisted below:

1. Rainfall and leaching
2. Acidic parent material
3. Organic matter decay
4. Harvest of high yielding crops.

The above causes of soil acidity are more easily understood when we consider that a soil is acid when there is an abundance of acidic cations (pronounced cat-eyeon), like hydrogen (H+) and aluminium (Al+++) present compared to the alkaline cations like calcium (Ca++), magnesium (Mg++), potassium (K+), and sodium (Na+).

Rainfall and Leaching

Excessive rainfall is an effective agent for removing basic cations over a long time period (thousands of years). In Oklahoma, for example, we can generally conclude that soils are naturally acidic if the rainfall is above 30 inches per year. Therefore, soils east of I-35 tend to be acidic and those west of I-35, alkaline.

There are many exceptions to this rule though, mostly as a result of item 4, intensive crop production. Rainfall is most effective in causing soils to become acidic if a lot of water moves through the soil rapidly. Sandy soils are often the first to become acidic because water perco-lates rapidly, and sandy soils contain only a small reservoir of bases (buffer capacity) due to low clay and organic matter contents. Since the effect of rainfall on acid soil development is very slow, it may take hundreds of years for new parent material to become acidic under high rainfall.

Parent Material

Due to differences in chemical composition of parent materials, soils will become acidic after different lengths of time. Thus, soils that developed from granite material are likely to be more acidic than soils developed from calcareous shale or limestone.

Organic Matter Decay

Decaying organic matter produces H+ which is responsible for acidity. The carbon dioxide (CO2) produced by decaying organic matter reacts with water in the soil to form a weak acid called carbonic acid. This is the same acid that develops when CO2 in the atmosphere reacts with rain to form acid rain naturally.

Several organic acids are also produced by decaying organic matter, but they are also weak acids. Like rainfall, the contribution to acid soil development by decaying organic matter is generally very small, and it would only be the accumulated effects of many years that might ever by measured in a field.

Crop Production

Harvesting of crops has its affect on soil acidity development because crops absorb the lime-like elements, as cations, for their nutrition.

When these crops are harvested and the yield is removed from the field, then some of the basic material responsible for counteracting the acidity developed by other processes is lost, and the net affect is increased soil acidity.

Increasing crop yields will cause greater amounts of basic material to be removed. Grain contains less basic materials than leaves or stems.

For this reason, soil acidity will develop faster under continuous wheat pasture than when grain only is harvested. High yielding forages, such as bermudagrass or alfalfa, can cause soil acidity to develop faster than with other crops.

The approximate amount of lime-like elements removed from the soil by a 30 bushel wheat crop. Note that there is almost four times as much lime material removed in the forage as the grain.

This explains why wheat pasture that is grazed out will become acidic much faster than when grain alone is produced. Using 50 percent ECCE lime, it would take about one ton every 10 years to maintain soil pH when straw (or forage) and grain are produced annually at the 30 bushel per acre level.

The use of fertilizers, especially those supplying nitrogen, has often been blamed as a cause of soil acidity. Although acidity is produced when ammonium containing materials are transformed to nitrate in the soil, this is countered by other reactions and the final crop removal of nitrogen in a form similar to that in the fertilizer.

What Happens in Acid Soils

Knowing the soil pH helps identify the kinds of chemical reactions that are likely to be taking place in the soil. Generally, the most important reactions from the standpoint of crop production are those dealing with solubilities of compounds or materials in soils. In this regard, we are most concerned with the affects of pH on the availability of toxic elements and nutrient elements.

Toxic elements like aluminium and manganese are the major causes for crop failure in acid soils. These elements are a problem in acid soils because they are more soluble at low pH. In other words, more of the solid form of these elements will dissolve in water when the pH is acid. There is always a lot of aluminium present in soils because it is a part of most clay particles.

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Infectious Plant Diseases and their Control

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A systems approach to plant health includes preventing and controlling diseases. There is one thing that must be kept in mind to be successful at plant disease control. You do not control a particular disease by doing any one particular thing. Diseases are managed, prevented or controlled by groups of practices loosely grouped under the concepts of the systems approach.

Infectious organisms, called plant pathogens, are part of the environment whenever and wherever plants are in existence. Surely, we do try and practice sanitation to limit their numbers. However, we cannot achieve total "sterility" of our greenhouses. We must manage the environments of our crops so plant pathogens cannot develop and flourish.

Cultivated plants are usually more susceptible to disease than their wild relatives, partly because large numbers of the same plant are often grown closely together in pure stands. Disease-causing organisms (pathogens) often get established under these conditions. Once this happens, they may spread rapidly. In addition, many of our valuable crop and ornamental plants are basically very susceptible to disease and would have difficulty surviving in undisturbed nature.

Finally, cultivation – be it geraniums, cabbages or oak trees – constantly disturbs nature and tends to create environmental stresses. Stresses often weaken plants and subject them to pathogenic infections.

What Causes a Plant Disease?

A plant can become diseased when it is continuously disturbed over a fairly long period of time by some factor or group of factors in its environment. Sometimes one of these disturbing factors is a living, infectious pathogen. Living pathogens include bacteria, fungi, viruses, nematodes and parasitic seed plants. Only if the disturbances are severe enough to produce noticed effects, called symptoms of poor health, do we call the plant diseased. Obviously, knowledge of normal growth habits, variety (cultivar) characteristics and normal variability of plants within a species is required for recognition of a diseased condition.

Protect cuttings from drying out before sticking to help control stem rots.

A System of Plant Disease Management

Employing health management systems is the most important way to proceed and succeed in growing plants profitably in a greenhouse or nursery. These are not complicated or difficult. They are simply based on the fact that a plant should not be chronically disturbed by elements of its environment with which it is has difficulty coping. Such plants are stressed and are more subject to the infection and development of infectious pathogens.

Reduction of plant stress involves awareness of environmental disturbances. It also involves trying to set up and maintain a reasonably balanced environment for each and every type of plant in the greenhouse or nursery.

Beyond Stress Management

As you have no doubt realized, when you are growing dozens of different kinds of plants, it is impossible to grow each one at its precisely balanced environment. Furthermore, there are some things you cannot control as you would like.

The amount of sunlight is probably the environmental factor that gives you the most trouble. Other things happen as a result of new workers tending to the plants differently, equipment breakdowns or changes in fertilizers, growing media, etc. In other words, a bumper sticker saying “Stress Happens” might be a good one for greenhouse growers!

Pathogen Management Practices

When stress happens, plant pathogens are primed to strike. We must turn our attention to methods of dealing with pathogens. Successful pathogen management is based on accurate diagnosis of the cause, thorough knowledge of the pathogen and its disease cycle, how the host and pathogen interact with various environmental factors, practicality of possible actions and cost of various practices.

The most important point in controlling a plant disease is choosing the best methods for a given situation. The best methods of control for one type of disease on a certain host may not be the best methods for another disease on the same or a different plant. Also, the use of several control measures (integrated control) is often needed. For example, several cultural control practices are often combined with a protective fungicide spray, sprench or drench program.

Infectious diseases in greenhouses or nurseries are controlled by one of four basic methods:

Exclusion. Disease-causing organisms should be excluded from production areas whenever possible. Any plant pathologist will have the practice of purchasing only pathogen-free plants, cuttings or seed high on his or her list. So what else is new? The problem is that you cannot be certain you have done this, even if you do pay extra money for “indexed” or “cultured” material. Try to do the best you can, of course.

Sanitation in and around the greenhouse is an excellent pathogen exclusion measure, as well. It involves cleaning and disinfecting potting benches, soil bins, head houses, greenhouse benches, tools and equipment. Disinfecting tools and equipment is a most valuable practice to help prevent the spread of many viruses and bacterial diseases. Growing crops on raised benches is an important way to keep pathogens away from crops. Using potting media that is relatively free of plant pathogens is yet another exclusion practice. Controlling weeds that harbour insects to spread pathogens or offer survival for pathogens between crops is important.

There are other exclusion practices that serve to prevent spread of pathogens from one area to another in the greenhouse or on any particular bed or bench. The most important of these is to prevent a great number of persons from handling stock plants. Another is to avoid splashing water from plant to plant whenever possible.

Protection. Growing resistant varieties is a form of protection. The genetics of the "host" protect it from pathogen infection or development. This is the most common form of protection from disease. After all, most plants do not get most diseases.

Most of us readily recognize that plants can be protected from pathogens by uniform and timely applications of disease-control chemicals (fungicides, bactericides and nematicides). Diseases caused by fungi are most commonly treated in this manner. Whereas there are many specific products that are particularly useful for particular fungal diseases, the trend of late is to use broad-spectrum products or combinations of products for general fungal disease treatment when necessary.

For leaf and flower diseases, a combination of mancozeb or chlorothalonil (contact protectants) and thiophanate-methyl (a systemic protectant) is very popular now. The new strobilurins are very useful. Flutolanil is coming into widespread use now.

The group of DMI systemics are useful for powdery mildews, rusts and other diseases. For root rots, a combination of thiophanate-methyl, fludioxinal or iprodione and etridiazole or mefenoxam is commonly used. When applying fungicides, it is important to have the fungicide on the plant surface or in the soil before infection takes place. Most fungicides are protectants, not eradicants. Prevention of pathogen infection and development conditions. Many of these practices come under the stress management practices discussed earlier. It never ceases to amaze me how tough plants are if they are not stressed.

Many commonly suggested cultural practices serve to manage infectious plant diseases through the idea of preventing spread, infection and development of pathogens. These generally involve altering the air and soil environment. Keeping plants spaced or growing plants on heated floors to allow foliage to dry quickly after watering are important examples. Spacing plants ensures adequate light to counter stress. It also favors uptake of needed nutrients for good plant health.

For root health, avoid overwatering and persistent wetness of the growing media. Also, avoid excessive dryness between waterings, use well-aerated and well-drained growing media and do not allow soluble salts to build up. Injured roots favour root-rotting pathogens.

Eradication. Plant pathogens can be eliminated (eradicated) and insects that spread pathogens can be controlled by pesticides. Removal and destruction of diseased plants is a common way to help control systemic bacterial and viral diseases. These plants may be the crop you are growing, weeds or other, less important host plants. Other eradication methods include crop rotation, which "starves" out soil-invading pathogens in nurseries.

As you can see from the above remarks, most of the disease-controlling systems we use today do not involve spraying or drenching with chemicals. These growing practices are effective and widely used by good growers. They take planning and follow through. This is truly a systems approach to plant health management.

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Chemical Process Methods in Industrial Distillation

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In a "scientific" sense, a chemical process is a method or means of somehow changing one or more chemicals or chemical compounds. Such a chemical process can occur by itself or be caused by somebody. Such a chemical process commonly involves a chemical reaction of some sort. In an "engineering" sense, a chemical process is a method intended to be used in manufacturing or on an industrial scale to change the composition of chemical(s) or material(s), usually using technology similar or related to that used in chemical plants or the chemical industry.

Neither of these definitions is exact in the sense that one can always tell definitively what is a chemical process and what is not; they are practical definitions. There is also significant overlap in these two definition variations. Because of the inexactness of the definition, chemists and other scientists use the term "chemical process" only in a general sense or in the engineering sense. However, in the "process (engineering)" sense, the term "chemical process" is used extensively. The rest of the article will cover the engineering type of chemical process.

Although this type of chemical process may sometimes involve only one step, often multiple steps, referred to as unit operations, are involved. In a plant, each of the unit operations commonly occur in individual vessels or sections of the plant called units. Often, one or more chemical reactions are involved, but other ways of changing chemical (or material) composition may be used, such as mixing or separation processes.

The process steps may be sequential in time or sequential in space along a stream of flowing or moving material. For a given amount of a feed (input) material or product (output) material, an expected amount of material can be determined at key steps in the process from empirical data and material balance calculations. These amounts can be scaled up or down to suit the desired capacity or operation of a particular chemical plant built for such a process.

More than one chemical plant may use the same chemical process, each plant perhaps at differently scaled capacities. In addition to chemical plants for producing chemicals, chemical processes with similar technology and equipment are also used in oil refining and other refineries, natural gas processing, polymer and pharmaceutical manufacturing, food processing, and water and wastewater treatment.

Process Integration

Process integration is a term in chemical engineering which has two possible meanings.

1. A holistic approach to process design which emphasizes the unity of the process and considers the interactions between different unit operations from the outset, rather than optimising them

separately. This can also be called *integrated process design* or *process synthesis*. El-Halwagi (1997 and 2006) and Smith (2005) describe the approach well. An important first step is often *product design* (Cussler and Moggridge 2003) which develops the specification for the product to fulfil its required purpose.

2. *Pinch analysis*, a technique for designing a process to minimise energy consumption and maximise heat recovery, also known as *heat integration*, *energy integration* or *pinch technology*. The technique calculates thermodynamically attainable *energy targets* for a given process and identifies how to achieve them. A key insight is the pinch temperature, which is the most constrained point in the process.

   The most detailed explanation of the techniques is by Linnhoff et al. (1982), Shenoy (1995) and Kemp (2006). This definition reflects the fact that the first major success for process integration was the thermal pinch analysis addressing energy problems and pioneered by Linnhoff and co-workers. Later, other pinch analyses were developed for several applications such as mass-exchange networks (El-Halwagi and Manousiouthakis, 1989), water minimisation (Wang and Smith, 1994), and material recycle (El-Halwagi et al., 2003). A very successful extension was "Hydrogen Pinch", which was applied to refinery hydrogen management. This allowed refiners to minimise the capital and operating costs of hydrogen supply to meet ever stricter environmental regulations and also increase hydrotreatre yields.

In the context of chemical engineering, Process Integration can be defined as a holistic approach to process design and optimisation, which exploits the interactions between different units in order to employ resources effectively and minimise costs.

Note that Process Integration is not limited to the design of new plants, but it also covers retrofit design (e.g. new units to be installed in an old plant) and the operation of existing systems. Nick Hallale (2001), in his article in Chemical Engineering Progress provided a state of the art review. He explained that process integration far wider scope and touches every area of process design. Industries are making more money from their raw materials and capital assets while becoming cleaner and more sustainable.

The main advantage of process integration is to (this is wrong) consider a system as a whole (i.e. integrated or holistic approach) in order to improve their design and/or operation. In contrast, an analytical approach would attempt to improve or optimise process units separately without necessarily taking advantage of potential interactions among them.

For instance, by using process integration techniques it might be possible to identify that a process can use the heat rejected by another unit and reduce the overall energy consumption, even if the units are not running at optimum conditions on their own. Such an opportunity would be missed with an analytical approach, as it would seek to optimise each unit, and thereafter it wouldn't be possible to re-use the heat internally.

Typically, process integration techniques are employed at the beginning of a project (e.g. a new plant or the improvement of an existing one) to screen out promising options to optimise the design and/or operation of a process plant.

Also it is often employed, in conjunction with simulation and mathematical optimisation tools to identify opportunities in order to better integrate a system (new or existing) and reduce capital and/or operating costs.

Most process integration techniques employ Pinch analysis or Pinch Tools to evaluate several processes as a whole system. Therefore, strictly speaking, both concepts are not the same, even if in certain contexts they are used interchangeably.

The review by Nick Hallale (2001) explains that in the future, several trends are to be expected in the field. In the future, it seems probable that the boundary between targets and design will be blurred and that these will be based on more structural information regarding the process network. Second, it is likely that we will see a much wider range of applications of process integration.

There is still much work to be carried out in the area of separation, not only in complex distillation systems, but also in mixed types of separation systems. This includes processes involving solids, such as flotation and crystallisation. The use of process integration techniques for reactor design has seen rapid progress, but is still in its early stages. Third, a new generation of software tools is expected. The emergence of commercial software for process integration is fundamental to its wider application in process design.

Adsorption

Adsorption is the adhesion of atoms, ions, biomolecules or molecules of gas, liquid, or dissolved solids to a surface. This process creates a film of the adsorbate (the molecules or atoms being accumulated) on the surface of the adsorbent. It differs from absorption, in which a fluid permeates or is dissolved by a liquid or solid. The term *sorption* encompasses both processes, while *desorption* is the reverse of adsorption. It is a *surface phenomenon*.

Similar to surface tension, adsorption is a consequence of surface energy. In a bulk material, all the bonding requirements (be they ionic, covalent, or metallic) of the constituent atoms of the material are filled by other atoms in the material. However, atoms on the surface of the adsorbent are not wholly surrounded by other adsorbent atoms and therefore can attract adsorbates. The exact nature of the bonding depends on the details of the species involved, but the adsorption process is generally classified as physisorption (characteristic of weak van der Waals forces) or chemisorption (characteristic of covalent bonding). It may also occur due to electrostatic attraction.

Adsorption is present in many natural physical, biological, and chemical systems, and is widely used in industrial applications such as activated charcoal, capturing and using waste heat to provide cold water for air conditioning and other process requirements (adsorption chillers), synthetic resins, increase storage capacity of carbide-derived carbons for tunable nanoporous carbon, and water purification. Adsorption, ion exchange, and chromatography are sorption processes in which certain adsorbates are selectively transferred from the fluid phase to the surface of insoluble, rigid particles suspended in a vessel or packed in a column. Lesser known, are the pharmaceutical industry applications as a means to prolong neurological exposure to specific drugs or parts thereof.

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Fundamental Concepts of Mechanics

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Space

Space is the boundless, three-dimensional extent in which objects and events occur and have relative position and direction. Physical space is often conceived in three linear dimensions, although modern physicists usually consider it, with time, to be part of a boundless four-dimensional continuum known as spacetime. In mathematics one examines "spaces" with different numbers of dimensions and with different underlying structures. The concept of space is considered to be of fundamental importance to an understanding of the physical universe although disagreement continues between philosophers over whether it is itself an entity, a relationship between entities, or part of a conceptual framework.

Debates concerning the nature, essence and the mode of existence of space date back to antiquity; namely, to treatises like the *Timaeus* of Plato, or Socrates in his reflections on what the Greeks called khora (i.e. "space"), or in the *Physics* of Aristotle (Book IV, Delta) in the definition of *topos* (i.e. place), or even in the later "geometrical conception of place" as "space *qua* extension" in the *Discourse on Place* (*Qawl fi al-Makan*) of the 11th century Arab polymath Alhazen. Many of these classical philosophical questions were discussed in the Renaissance and then reformulated in the 17th century, particularly during the early development of classical mechanics. In Isaac Newton's view, space was absolute - in the sense that it existed permanently and independently of whether there were any matter in the space. Other natural philosophers, notably Gottfried Leibniz, thought instead that space was a collection of relations between objects, given by their distance and direction from one another.

In the 18th century, the philosopher and theologian George Berkeley attempted to refute the "visibility of spatial depth" in his *Essay Towards a New Theory of Vision*. Later, the metaphysician Immanuel Kant said neither space nor time can be empirically perceived, they are elements of a systematic framework that humans use to structure all experiences. Kant referred to "space" in his *Critique of Pure Reason* as being: a subjective "pure *a priori* form of intuition", hence it is an unavoidable contribution of our human faculties.

In the 19th and 20th centuries mathematicians began to examine non-Euclidean geometries, in which space can be said to be *curved*, rather than *flat*. According to Albert Einstein's theory of general relativity, space around gravitational fields deviates from Euclidean space. Experimental tests of general relativity have confirmed that non-Euclidean space provides a better model for the shape of space.

Time

Time is a part of the measuring system used to sequence events, to compare the durations of events and the intervals between them, and to quantify rates of change such as the motions of objects. The temporal position of events with respect to the transitory present is continually changing; events happen, then are located further and further in the past. Time has been a major subject of religion, philosophy, and science, but defining it in a non-controversial manner applicable to all fields of study has consistently eluded the greatest scholars. A simple definition states that "time is what clocks measure".

Figure: ***The flow of sand in an hourglass can be used to keep track of elapsed time. It also concretely represents the present as being between the past and the future.***

Time is one of the seven fundamental physical quantities in the International System of Units. Time is used to define other quantities — such as velocity — so defining time in terms of such quantities would result in circularity of definition. An operational definition of time, wherein one says that observing a certain number of repetitions of one or another standard cyclical event (such as the passage of a free-swinging pendulum) constitutes one standard unit such as the second, is highly useful in the conduct of both advanced experiments and everyday affairs of life.

The operational definition leaves aside the question whether there is something called time, apart from the counting activity just mentioned, that flows and that can be measured. Investigations of a single continuum called spacetime bring questions about space into questions about time, questions that have their roots in the works of early students of natural philosophy.

Two contrasting viewpoints on time divide many prominent philosophers. One view is that time is part of the fundamental structure of the universe, a dimension in which events occur in sequence. Sir Isaac Newton subscribed to this realist view, and hence it is sometimes referred to as Newtonian time. Time travel, in this view, becomes a possibility as other "times" persist like frames of a film strip, spread out across the time line.

The opposing view is that *time* does not refer to any kind of "container" that events and objects "move through", nor to any entity that "flows", but that it is instead part of a fundamental intellectual structure (together with space and number) within which humans sequence and compare events. This second view, in the tradition of Gottfried Leibniz and Immanuel Kant, holds that *time* is neither an event nor a thing, and thus is not itself measurable nor can it be travelled.

Temporal measurement has occupied scientists and technologists, and was a prime motivation in navigation and astronomy. Periodic events and periodic motion have long served as standards for units of time. Examples include the apparent motion of the sun across the sky, the phases of the moon, the swing of a pendulum, and the beat of a heart. Currently, the international unit of time, the second, is defined in terms of radiation emitted by caesium atoms. Time is also of significant social importance, having economic value ("time is money") as well as personal value, due to an awareness of the limited time in each day and in human life spans.

Ray Cummings, an early writer of science fiction, wrote in 1922, "Time... is what keeps everything from happening at once", a sentence repeated by scientists such as C. J. Overbeck, and John Archibald Wheeler.

Velocity

In physics, velocity is speed in a given direction. Speed describes only how fast an object is moving, whereas velocity gives both the speed and direction of the object's motion. To have a constant velocity, an object must have a constant speed and motion in a constant direction. Constant direction, typically constrains the object to motion in a straight path. A car moving at a constant 20 kilometres per hour in a circular path does not have a constant velocity. The rate of change in velocity is acceleration. Velocity is a vector physical quantity; both magnitude and direction are required to define it. Thescalar absolute value (magnitude) of velocity is speed, a quantity that is measured in metres per second (m/s or ms^{-1}) when using the SI (metric) system.

For example, -5 metres per second" is a scalar and not a vector, whereas -5 metres per second east" is a vector. The *average* velocity v of an object moving through a displacement (Δx) during a time interval (Δt) is described by the formula:

$$\overline{v} = \frac{\Delta x}{\Delta t}.$$

The rate of change of velocity is acceleration – how an object's speed or direction of travel changes over time, and how it is changing at a particular point in time.

Equation of Motion

The velocity vector v of an object that has positions $x(t)$ at time t and $x(t + \Delta t)$ at time $t + \Delta t$, can be computed as the derivative of position:

$$v = \lim_{\Delta t \to 0} \frac{x(t + \Delta t) - x(t)}{\Delta t} = \frac{dx}{dt}.$$

Average velocity magnitudes always smaller than or equal to average speed of a given particle. Instantaneous velocity is always tangential to trajectory. Slope of tangent of position or displacement time graph is instantaneous velocity and its slope of chord is average velocity.

The equation for an object's velocity can be obtained mathematically by evaluating the integral of the equation for its acceleration beginning from some initial period time t_0 to some point in time later t_n. The final velocity v of an object which starts with velocity u and then accelerates at constant acceleration a for a period of time Δt is:

$$v = u + a\Delta t.$$

The average velocity of an object undergoing constant acceleration is $\frac{(u+v)}{2}$, where u is the initial velocity and v is the final velocity. To find the position, x, of such an accelerating object during a time interval, Δt, then:

$$\Delta x = \frac{(u + v)}{2}\Delta t.$$

When only the object's initial velocity is known, the expression,

$$\Delta x = u\Delta t + \frac{1}{2}a\Delta t^2,$$

can be used.

This can be expanded to give the position at any time t in the following way:

$$x(t) = x(0) + \Delta x = x(0) + u\Delta t + \frac{1}{2}a\Delta t^2,$$

These basic equations for final velocity and position can be combined to form an equation that is independent of time, also known as Torricelli's equation:

$$v^2 = u^2 + 2a\Delta x.$$

The above equations are valid for both Newtonian mechanics and special relativity. Where Newtonian mechanics and special relativity differ is in how different observers would describe the same situation. In particular, in Newtonian mechanics, all observers agree on the value of *t* and the transformation rules for position create a situation in which all non-accelerating observers would describe the acceleration of an object with the same values. Neither is true for special relativity. In other words only relative velocity can be calculated.

In Newtonian mechanics, the kinetic energy (energy of motion), E_K, of a moving object is linear with both its mass and the square of its velocity:

$$E_K = \frac{1}{2}mv^2.$$

The kinetic energy is a scalar quantity.

Escape velocity is the minimum velocity a body must have in order to escape from the gravitational field of the earth. To escape from the Earth's gravitational field an object must have greater kinetic energy than its gravitational potential energy. The value of the escape velocity from the Earth's surface is approximately 11100 m/s.

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Applied Pig Breeding for Genetical Goal

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Introduction

Pigs are used mainly for producing human foods. Meat cuts are the main interest, but other products derived from the carcass, such as legs and noses (e.g. for Chinese market), are used for human consumption. Secondary uses of pigs include manure production and the fulfilment of cultural needs. In medical research, pigs are also used as models of humans. Pigs are kept in a broad spectrum of production environments around the world, but in Denmark the vast majority are kept in intensive housing conditions with a controlled climate; a minority of Danish pigs are kept outside in free range environments.

Denmark is among the world's largest pig producers. In 2009, 19.3 million pigs were slaughtered in Denmark, which corresponds to 2 million tonnes of meat. Worldwide, about 93 million tonnes of pig meat was generated by slaughter in 2009. In Denmark, 94% of the meat produced in 2009 was exported; and Germany (30% of that meat), United Kingdom (15%), Japan (7%) and China (7%) were among the larger importers of Danish pig meat (Landbrug og Fødevarer, 2010).

Artificial insemination (AI) with fresh (non-frozen) semen is used in most matings. Boars can produce about 50 doses of semen per week, and this allows them to be intensively selected. Purebred Landrace, Yorkshire and Duroc sows farrow 15.3, 15.3 and 9.8 piglets per litter on average. Gilts reach sexual maturity at 6–7 months of age, and their average gestation length is 116 days.

Breeds

Danish pig production is based mainly on three breeds: Duroc, Landrace and Yorkshire. Duroc is used as a terminal sire on Landrace x Yorkshire (LY) sows to produce crossbred pigs for Danish production herds. Other countries use breeds with the same names and similar origin as these 'Danish' breeds, but their populations differ as the result of, among things, different breeding goals and the restricted exchange of genetic material.

Hampshire, Piétrain and Berkshire are also used in some countries, and locally other breeds continue to have some commercial influence. China, the world's largest swine industry, has been based on roughly six types of pig, defined by geographical location and origin.

However, a rapid transition is taking place in China to US and/or European breeds, and now Piétrain, Duroc, Landrace and Yorkshire are the most commonly used breeds in modern cross-breeding systems. Durocs were imported from North America to Denmark in the late 1970s. Besides its high growth capacity, good carcass traits and high feed efficiency, the breed is recognized by its red-brown colour. Yorkshire and Landrace are both white. They are known for their maternal qualities (i.e. they have large litters and nurse their piglets well).

Breeding Goal

The breeding goal is to breed pigs that will generate the highest possible economic return for commercial pig producers over the coming 5–10 years. This breeding goal is decided by commercial pig producers with the guidance of the Breeding & Genetics section at the Danish Agricultural and Food Council. Economic values for most traits are based on a bioeconomic model. This model simulates incomes and costs of each trait in a 'future' production herd; it can be amended to reflect political concerns. The breeding goal is different for paternal (Duroc) and maternal (Landrace, Yorkshire) breeds.

Genetic Evaluation and Parameters

Multiple-trait animal models are used in the genetic evaluation of groups of 2–4 traits. For instance, estimated breeding values (EBVs) for feed efficiency, the two growth traits and lean meat percentage are calculated using a four-trait model. Although genetic correlations are relatively small this is especially advantageous for feed efficiency, because animals without records on feed efficiency, but with records on one or more of the other traits, obtain EBVs that are based on correlated information.

The explanatory effects used in the genetic evaluations to account for environmental effects differ from trait to trait. Typical effects are sex, herd-year-month of registration, common environment for litters, common environment effect for the housing group of pigs, and weight of the animal at the onset of the registration period (e.g. growth 30–100 kg). The bivariate model for number of piglets alive after day 5 and litter size also includes effects of parity of sow, the sow's age at 1st farrowing (1st parity only), farrowing interval (later parities only) and type of fertilization (AI or natural). The parameters used in genetic evaluation and in the breeding programme for Landrace pigs are summarize.

The heritabilities and correlations are similar for Duroc and Yorkshire, whereas variances in some traits differ. Strength of legs and claws, number of pigs alive after day 5, and sow longevity have low heritability (0.08–0.17). The last two traits are not evaluated for Duroc. Genetic correlations between male (e.g. growth, feed efficiency) and female traits (e.g. no. piglets alive after day 5, sow longevity) tracked in the Danish system are not estimated. Research on foreign pig populations suggests that the genetic correlations between growth and reproductive traits are either unfavourable (e.g. Holm et al., 2004) or close to zero (e.g. Arango et al., 2005).

Organization and Breeding Programme

Danish pig breeding is organized around a classical breeding pyramid. In 2010 the Danish pig population consisted of 32 breeding herds (1785, 2210 and 2717 Duroc, Yorkshire and Landrace sows, respectively), 153 multiplier herds (69 700 purebred sows) and 2601 production herds with 1.1 million crossbred sows.

There is some overlap between the figures, as 29 breeding herds are also multiplier herds. The breeding herds form a closed nucleus with no imports from lower tiers in the pyramid or foreign populations.

Thus it is only selection and mating decisions made in the breeding herds that influence the additive genetic trends in the population. The current average genetic level in production herds corresponds approximately to the average genetic level observed in the breeding herds 1–2 generations ago. (Transmission of genes takes 1 and 2–3 generations for boars and sows, respectively.)

Breeders send their best boars to AI-stations and also sell approximately 1000 (mainly Duroc) boars per year to production herds. Purebred females are sold to multiplier herds and, in some cases, directly to production herds. Hence, breeders successfully breeding superior pigs earn more money than their less successful competitors. This is an important motivation for breeders to do their best when they record breeding goal traits, selecting animals with the best EBVs and ensuring optimal matings.

The main function of multiplier herds is to facilitate the transmission of genetic progress made in breeding herds to production herds. In practice, this means producing crossbred females (LY) that are sold to production herds. Multiplier herds receive purebred Landrace and Yorkshire females from breeding herds. Breeding decisions in production herds are not relevant to future generations of the pig population. Such herds exist primarily for the production of pigs for slaughter. As the vast majority of pigs are raised in production herds, the breeding goal should reflect the circumstances in production herds, and ideally performance measures of breeding animals should be carried out in similar production environments.

Most traits are recorded in the purebred breeding herds. However, feed efficiency is recorded at the test station 'Bøgildgaard' and not in individual herds. The number of piglets alive per litter is recorded in multiplier herds as well as in breeding herds to provide sufficient accuracy of breeding values (trait only expressed by sows and low heritability). Slaughter loss is only recorded for slaughtered animals, which makes it impossible to have own records on active breeding animals. The remaining traits are recorded on most pigs in breeding herds — only approximately 25% of the pigs do not have their performance recorded, and this is due mainly to death, disease or experimental discrepancies.

The approximate proportion of tested pigs that are used for pure-breeding. Selection intensities are substantially higher for boars than gilts as a consequence of AI being used. These intensities are lower for Duroc as compared with the maternal breeds as a result of Duroc's smaller average litter size, smaller population, and because some Duroc boars are used for both breeding and production herds. The use of selected boars varies substantially (i.e. the number of matings per boar ranges from 1–60).

Inbreeding only concerns breeding herds (i.e. purebred pigs) and is controlled by imposing an upper limit of 50–60 matings for a single boar, depending on the breed. Furthermore, a maximum of 40 half-and 2 full-brothers are accepted at the 'Bøgildgaard' test station for Yorkshire and Landrace, whereas a maximum of 100 half-and 3 full-brothers are accepted for Durocs. Breeders decide which matings to arrange on the basis of these limitations. Limiting the use of each boar is easy enough in practice, but it is not an optimal way of controlling inbreeding since it does not account for relationships among boars and their breeding values. Therefore, The Danish Agricultural & Food Council's Pig Research Centre is working on implementing optimum contribution selection of boars (Bendtsen, 2008).

Genomic EBVs based on a 62K SNP chip are currently being developed for all evaluated traits. They are expected to have the greatest impact on longevity, litter size and feed efficiency, where accurate EBVs are not available for young selection candidates. Conversely, they are expected to have little impact on the remaining evaluated traits. Potentially, genomic EBVs will also be developed for health traits that are not being evaluated today. Furthermore, genomic EBVs permit the collection of data on crossbred sows — and the subsequent use of this information, in connection with purebred animals. This helps to overcome problems with genotype by environment interactions and gene expression differences between pure and crossbred pigs due to different background genetics.

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Mammalian Cloning Methods and Applicatons

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Cloning is commonly perceived as a means of generating genetically identical individuals, but it can also be used to obtain genetically matched embryo-derived stem cells, which could potentially be used in the treatment of patients. A recent report offers the first 'proof of principle' of such cloning for therapeutic purposes, referred to as nuclear transplantation to produce stem cells for autologous transplantation.

Cloning is a mode of asexual reproduction in which all offspring have an identical nuclear genome to that of the parent. In recent years, mammalian cloning has been achieved by the introduction of somatic cell nuclei into fertilized eggs from which the zygotic nucleus has been removed. In our anthropocentric society, the recent success in animal cloning and its implications for humanity have captured the public's attention and imagination. But even though cloning has now been accomplished in several mammalian species, there are often severe complications associated with the procedure, and cloned animals are never quite the same as their parent.

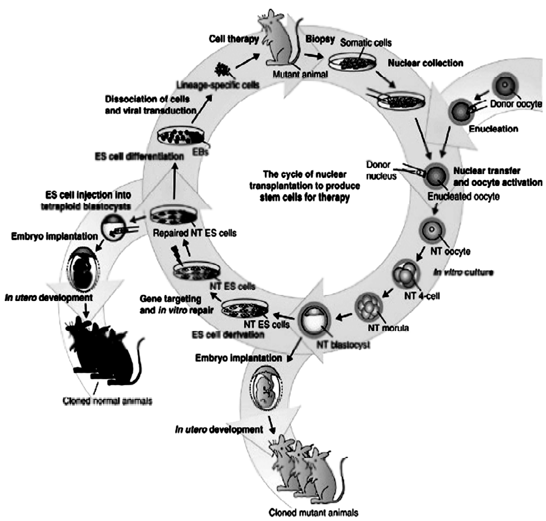
For instance, cloned embryos often exhibit developmental abnormalities, usually including excessive growth, referred to as large-offspring syndrome (LOS); in some cases, epigenetic aberrations have been reported, such as inappropriate X chromosome inactivation in cloned bovine fetuses and placentae.

Thus, only a very small proportion (less than 1%) of cloned mammals make it to birth. Many of the offspring that are born suffer from various defects, including obesity and liver and immunological defects; their chromosomes often have telomeres with variable lengths, possibly correlating with the donor cell type used for generating clones. Either individually or in combination, these symptoms may drastically shorten the lifespan of clones. It is issues such as these that have raised considerable concern about the cloning procedure and highlighted our lack of understanding of the basic biology of cloning.

Nuclear Transfer and Stem Cells

The technique of vertebrate somatic-cell nuclear transfer (also referred to as nuclear transplantation) was first developed half a century ago in amphibians, and the first cloned adult amphibians were described a decade later. Only in the last five years has the technique been used successfully for the production of viable cloned mammals. There are currently two elegantly simple protocols for the cloning of mammals by nuclear transfer.

The first relies on the fusion of a somatic cell and an enucleated egg and has been used to clone sheep, mice, goats, cows and pigs, whereas the second is based on nuclear microinjection and has been extensively used to generate cloned mice and also cloned pigs and goats. Both protocols involve the removal of the nucleus from an unfertilized egg (an oocyte) and its replacement with a nucleus from an adult cell or a cultured cell line; both rely on the premise that the microenvironment of the host oocyte - presumably its cytoplasm - can re-instruct the donor nucleus to adopt the behaviour of the removed oocyte nucleus. Thus, the donor nucleus is reprogrammed so that it becomes developmentally versatile (totipotent) and able to direct and execute the embryonic developmental program.



Cloning entire individuals by nuclear transfer is not the aim of most studies at present, however. Of far more interest is the potential to produce stem cells and, by combining the production of stem cells with nuclear transfer, to produce autologous stem cells that match the donor of the adult nucleus. Stem cells are cells that have the unique dual capacity for self-renewal and differentiation; in other words, they can not only divide to give identical stem-cell progeny, they can also differentiate into a wide variety of other cell types.

There are several categories of stem cell, including embryo-derived and lineage-specific stem cells: the former usually have a broader repertoire for differentiation than the latter. Stem cells isolated from the inner cell mass of the blastocyststage preimplantation mammalian embryo, known as embryonic stem or ES cells, can contribute to most but not all lineages; this pluripotency mirrors that of the inner cell mass. If included in embryos derived from more than one fertilized egg (chimeras), ES cells can contribute to the fetus itself (including the germ line) and extraembryonic mesoderm. ES cells can also be maintained as permanent, undifferentiated cell lines *in vitro* while still preserving their developmental potential. It has been two decades since ES cells were first isolated, and they remain the mainstay of mouse genome engineering because their genes can easily be manipulated *in vitro* - even down to individual base pairs - by standard gene-targeting and transgenesis techniques, while their developmental potential is retained. Genome modifications introduced into ES cells *in vitro* can be reintroduced into mice via inclusion in chimeric embryos. ES cells can also be induced to differentiate into defined-lineage cell types under appropriate conditions *in vitro.*

Embryonic stem cells are not the only type of stem cells - nor are they necessarily the most appropriate type for therapeutic purposes. Lineage-specific stem cells are the progenitors of specific differentiated cell lineages and are present in later-stage embryos and adults in organs such as skin, intestine, brain, and bone marrow. One issue that arises from the presence of such cells is that in animal cloning studies, donor nuclei have been taken from ostensibly differentiated somatic cells, but most donor cell populations are probably heterogenous, and it is not clear whether it is differentiated cells or rare lineage-specific stem cells in the population that give rise to clones. If the latter is the case, it may be that the rarity of stem cells leads to the low efficiencies of cloning - lower than when ES cells are used as nuclear donors. Protocols that can distinguish stem cells from differentiated cells would then need to be developed in order to increase overall efficiencies. A key question that has therefore persisted, and remained unanswered until recently, is whether highly specialised lineage-specific cells can be reprogrammed such that they can adopt a totipotent state, with the potential to differentiate into all possible cell types, and thus direct the developmental program used to generate a complete individual.

Reprogramming Differentiated Cells

Rudolph Jaenisch and colleagues, who have been at the forefront of nuclear-transfer work in mice, designed an experiment to address the issue of reprogramming differentiated lineage-specific cells. They chose lymphocytes as nuclear donors, as these are one of the few cell types of adult mammals whose genome is irreversibly changed as they mature, thereby making them genetically distinct and recognisable. B and T cells are the two classes of mature lymphocytes, expressing immunoglobulins (antibodies) and T-cell receptors, respectively. The type of antibody or T-cell receptor expressed is dictated by the rearrangement of each cell's genomic DNA; mature lymphocytes express only one specific antibody or receptor. Thus, in clones generated from B or T cells, the signature genomic rearrangement present in each donor cell nucleus would be preserved in all the cells of the cloned progeny.

Jaenisch and colleagues' study investigating the developmental potential and reprogramming of lymphocyte nuclei combined the technologies of mammalian cloning and ES cells in a two-step procedure that improved the efficiency of generating clones. First, they generated nuclear-transfer (NT) embryos by transfer of lymphocyte nuclei, but instead of re-implanting the embryos directly into the uteri of foster mothers, they used NT blastocysts to derive NT ES cells. They then took advantage of the 'tetraploid complementation' technique and injected their NT ES cells into tetraploid host blastocysts. Tetraploid cells preferentially form the extraembryonic tissues trophoblast and extraembryonic endoderm and are excluded from fetal tissues and extraembryonic mesoderm, whereas ES cells exclusively form the latter two tissues.

These experiments resulted in the production of cloned mice from adult lymphocyte nuclei, as could be recognised by the signature genomic rearrangements of lymphocytes. Jaenisch and colleagues have thus answered the previously unresolved question of whether terminally differentiated cells can provide nuclei for the production of clones, by demonstrating that at least some specialised nuclei can be reprogrammed. Perhaps the tetraploid extraembryonic component used in this procedure may be pivotal in helping overcome some of the defects that have otherwise consistently been observed in cloned embryos (such as enlargement of the placenta). Additionally, the nature of this experimental setup is less demanding of the NT cells, as they do not need to be truly totipotent because all the trophoblast and extraembryonic endoderm derivatives are derived from the tetraploid cells. The NT cells therefore need only to achieve pluripotency to generate the fetus. Also, the extended period of time in culture inherent in the ES-cell-derivation procedure may allow further or more complete reprogramming of the differentiated donor nuclei, ultimately leading to increased developmental potential.

The potential of the repaired NT ES cells was then tested in two ways. First, repaired NT ES cells were injected into tetraploid blastocysts for the generation of offspring. Normal embryos developed to birth from the tetraploid chimeras, indicating that the NT ES cells with the repaired *Rag2* gene retained their pluripotency. Furthermore, the presence of normal T and B cells in these mice proved that the repaired *Rag2* allele was functional. Second, repaired NT ES cells were differentiated in culture into hematopoietic stem cells (which form blood and immune cells), and the latter cells were transplanted into adult *Rag2* mutant mice. (Incorporation of the cells into the immune system was not entirely successful because of an immune barrier peculiar to the Rag2-deficient recipients, but this barrier was partially overcome by further manipulation of the immune system of the recipients.) Thus, the procedure was successful in restoring a modest degree of immune function in the mutant mice, but the difficulties encountered suggest that even genetically matched cells derived by nuclear transplantation may still face barriers to effective transplantation in some situations.

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Essential Component of the Modelling Process in Chemistry

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Currently, the role of argumentation is accepted as the focus point when constructing explanations, models, and theories, just as scientists use argumentation in order to connect their defended hypotheses with data or initial starting points. On the other hand, science teaching based on models or designed according to a modelling process uses argumentation as an essential tool for constructing theoretical and practical meanings. Sardα and Sanmarti (2000), among others, have developed different proposals along these lines, at the school level, with favourable results when introducing activities that promote discussion in class. From this standpoint, we develop the peculiarities in a chemistry course where trained teachers work with a given pattern that will help them to develop an argumentation and explanation about a chemical phenomenon. The results allow us to compare the argumentation created and to identify the possible differences due to the curricular planning and design that might be generating discussion in different grades.

Carrying out a scientific school activity is also modelling when that activity mixes experimentation and regulated discussion to promote a rational reconstruction of a phenomenon.

The Structure of the Modelling Process

The modelling process within the teaching process is complex, as the students might not understand the theories, their applications, or the specific language used - on the other hand, they may know the specific language of the theory but do not know how to apply it.

An expert (left) contrasts phenomenon with a theoretical model already known, or that the expert intuitively knows is going to work. Based on this model, the results of a certain experimental intervention are anticipated; if the expectation of the model coincides with the result of the experiment, the new phenomenon will be explained using the model's own theoretical bodies (entities). A beginner being introduced to the discipline needs to get used to the new culture, and must learn to ask himself or herself about quantities and relations. And he or she should also be able to intervene, experimentally using new instruments that require new ways of acting and understand through models still unknown. The beginner does not have the autonomy to act and needs the teacher's guidance during the scientific activities.

In the second case (right), the first thing that needs to be done is to present a certain group of facts (the ones that are going to be interpreted in a similar way), after the suggestion of a model that will allow establishing relations between those facts. These are more abstract than the ones that give a more detailed explanation, but are better rooted in the scientific culture and the student's cognition. Modelling is a process that occurs when students are learning to understand the facts they are observing. This process is continuous while building up connections and explanations, and becomes gets more and more complex.

To give content to the modelling idea, a chemistry course has been designed for future primary school teachers in the Faculty of Education in the Autonomous University of Barcelona (UAB). Using this approach, we start each theme with a story which gives context to the concepts to be learned and the experiments that will take place in class. This is a different approach to that of writing formulas without understanding them or repeating what is written in textbooks. The design of the course emphasizes the relation between Language-Model Experiment, and the activity in chemistry, according to the aims of the research group LIEC at the UAB.

This course was based on the above described and other modelling experiences. We outline in advance some general characteristics of the process. This is analysed in more detail and contrasted with new examples. First, we describe how the different learning processes can be developed. This is further outlined in the didactic units we designed:

Phenomenon: Burning iron wool and oxidising iron wool. Here, we take advantage of previous experiments on burning different materials and the oxidation of metals.

Critical Argumentation

The importance of argumentation in the construction of scientific meaning and its status as a fundamental characteristic of the scientific activity is accepted. Argumentation plays a key role in the construction of explanations, models, and theories because scientists use those arguments to connect data or other starting points. The construction of argumentation has been a challenge in science teaching because it is where practical and theoretical meanings are built. There exists several different proposals for working with argumentation in class. These proposals inspired the creation of a template. The argumentative node corresponds with ideas that permit one to solve doubts through pros and cons which give rise to a stronger conclusion.

Why Use a Template? From the ICT implementation point of view Pea (1993) highlights the importance of having access to tools for thinking more than for a solitary understanding. It was important to find tools which represent a material anchor for cognition. ICT tools and computer programmes can be considered as facilitating distributed cognition (everything that is thought, known, and done in the group is knowledge that is available for all its members) and it encourages cooperative work. In order to encourage the argumentation processes, the inclusion of an ARGU (Microsoft Excel) e-template was considered to assist in constructing the argument process, as a specific contribution which takes advantage of communication technologies to mediate or facilitate the learning process.

Placing emphasis on this type of tool allowed different results when compared to the ones used to develop individual skills. Therefore, the conception of educational goals is changed, going from individual control to the common action, through agreements and discussion, including disagreement. The idea of mediating the activity with artefacts involves the distribution of cognition between the individuals, the environment, and the mediator. It also involves a change to the basis of the activity that is being mediated with these artefacts.

In our research, we analysed the results obtained when using a guide for argumentation in a teaching and learning chemistry environment. This was experimental according to a semantic concept of a scientific theory which prioritized the construction in class of scientific facts, where theoretical statements reach meaning and through a modelling process - where argumentation is essential and is required.

Work Plan and Research Methodology

To help in the creation of the arguments, it was necessary to generate situations in which there could be reasonable doubt with regard to the results and different possible interpretations - both would make it necessary to look for convincing reasons concerning the explanation being given. As the students were not trained in argumentation in science classrooms, they would feel overwhelmed when required to do so. Because of this, a proposal was brought forward for students to work in a cooperative way with an Excel sheet. With this the students could interact and could be helped, not only during their first steps (giving them a text structure that should help as an argument pattern), but also in their final interpretations, when they are able to modify the given structure and to generate their own connectors.

The template contains a group of mediation questions that facilitate the formulation of the necessary ideas to give arguments about a phenomenon. At the same time, these questions show that the phenomenon (or fact) can be seen in different ways and thus generate doubt justification requirements. The class activity consisted in the reading of a text prepared by the teacher to introduce an experimental situation that the students must resolve.

Once the experiment has been completed, they must explain it, taking into account the errors, anomalies, and doubts that have been raised or that could be formulated by an outside observer who does not know anything about chemistry.

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Thermoluminescence Dosimetry Study of Dolomite and Calcium Fluorite

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Abstract

The thermoluminescence (TL) dosimetric properties of natural dolomite and calcium fluoride (CaF,) minerals collected from the Chhotaudepur district of Gujarat were investigated using beta irradiation from a Sr-90 source. TL growth and decay characteristics were studied for the as-received and thermally treated samples. Distinct glow peaks with dose-dependent intensity enhancement and peak shifting were observed. Fading studies revealed the rapid decay of low-temperature peaks, whereas high-temperature peaks showed better stability. The results indicate that dolomite and CaF, possess promising TL characteristics that are suitable for radiation dosimetry applications.

Keywords: Thermoluminescence dosimetry; Dolomite; Calcium fluoride; TL growth; TL fading

1. Introduction

In the present scientific world, ionizing radiation plays a vital role in engineering, medicine, science, and technology. Accurate determination of the absorbed radiation dose is essential for achieving the desired outcomes in these applications, as it enables a reliable assessment of energy deposition and its distribution within the exposed materials. Radiation dosimetry addresses this requirement through systems known as dosimeters, which measure the absorbed dose with high precision.

Thermoluminescence (TL) dosimetry is one of the most widely used techniques for measuring radiation dose. In TL dosimetry, the relationship between the thermoluminescence signal and the absorbed dose must be established through appropriate calibration. With advancements in thermoluminescent materials and readout instrumentation, Thermoluminescence Dosimeters (TLDs) have found extensive applications in personal monitoring, environmental radiation surveillance, and clinical and medical dosimetry. Their high sensitivity, small size, wide dose range, reusability, and relative insensitivity to environmental conditions make them superior to conventional film badge techniques, which were used in earlier decades.

An ideal TL phosphor should possess a high concentration of stable trapping centers, high emission efficiency, suitable trap depth with minimal fading, and a stable crystalline lattice that is unaffected by radiation exposure. Parameters such as dose response, glow curve characteristics, sensitivity, superlinearity, and fading behavior are crucial for evaluating the suitability of TL materials for dosimetric applications.

The present work focuses on the thermoluminescence dosimetric study of natural dolomite and calcium fluoride (CaF,) minerals collected from the Chhotaudepur district of Gujarat. The investigation was divided into two parts: (i) TL growth studies and (ii) TL decay studies, with the aim of assessing their potential use as effective TL dosimetric materials.

2. Experimental

Natural mineral samples of dolomite and calcium fluoride (CaF,) were collected from Chhotaudepur and Amba Dungar regions, respectively, located in the Chhota Udaipur district of Gujarat, India. The collected samples were thoroughly cleaned using distilled water to remove adhering soil and dust particles and then dried at ambient temperature. The dried samples were finely ground using an agate mortar and pestle to obtain uniform grain size suitable for thermoluminescence (TL) measurements.

Thermoluminescence measurements were carried out using a microcontroller-based Nucleonix TL reader. For each TL measurement, an equal mass of approximately 5 mg of the powdered sample was used to ensure consistency and reproducibility of results.

The experimental work was divided into two parts: TL growth studies and TL decay studies. In the first part, comparative TL growth behavior and glow curve characteristics were investigated for both as-received samples and samples subjected to thermal treatment. Annealing was performed at 800 °C followed by rapid quenching to room temperature. After thermal treatment, the samples were irradiated with different beta radiation doses of 25 Gy, 50 Gy, 100 Gy, 150 Gy, and 250 Gy using a Sr-90 beta source. The resulting TL glow curves were recorded, and the corresponding peak temperatures and peak intensities were analyzed and tabulated for comparative study.

The second part of the investigation focused on TL decay (fading) characteristics. Both as-received and annealed–quenched samples were irradiated with a fixed beta dose of 25 Gy from the Sr-90 source. TL measurements were recorded immediately after irradiation and subsequently after storage periods of 24 h, 48 h, 100 h, 150 h, and 280 h at room temperature. The variation in TL peak intensity and peak temperature with storage time was studied to assess the stability and fading behavior of the dosimetric signals. Relevant data were tabulated to facilitate comparison and interpretation.

3. Results and discussions

3.1 Thermoluminescence Growth

The thermoluminescence (TL) glow curves of dolomite and calcium fluoride (CaF,) samples exhibit distinct and well-resolved peaks, indicating multiple trapping centers with different thermal stabilities.

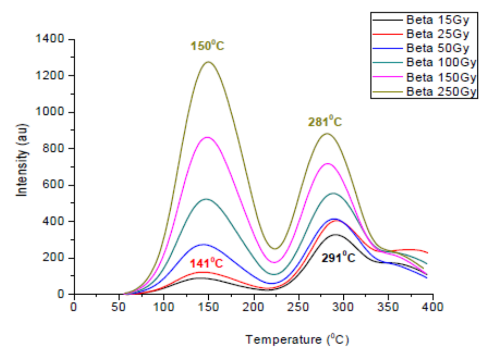

Fig.-1: TL glow curves for dolomite exposed to various amount of Sr-90 beta doses.

Table-1

Sr. No.	Beta dose (Gy)	Peak Temp. (^{0}C)	Peak Intensity (A.U.)
1.	15	141, 291	89, 328
2.	25	144, 292	123, 405
3.	50	145, 290	274, 414
4.	100	147, 288	523, 554
5.	150	149, 282	862, 717
6.	250	150, 281	1275, 882

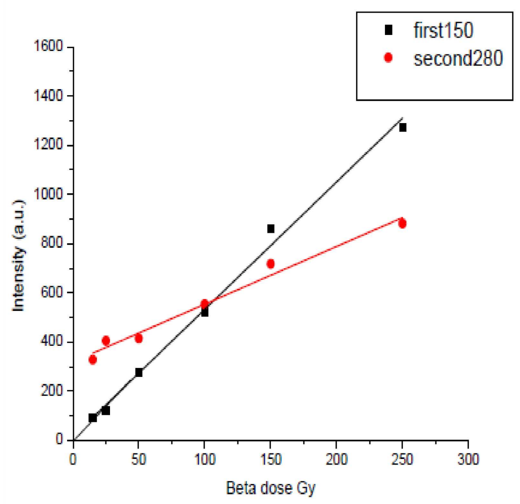

Fig.-2: Dose response curve of dolomite

For the dolomite sample, the TL glow curve displays two isolated peaks at approximately 141/ °C and 291/ °C. With increasing beta radiation dose, the first peak (141/ °C) shifts towards higher temperatures, reaching ~150/ °C, while the second peak (291/ °C) shifts towards lower temperatures, moving to ~281/ °C. Simultaneously, the intensities of both peaks increase progressively with dose, indicating enhanced population of the traps. The variation of TL peak temperature and intensity with respect to applied dose is summarized in Table 1.

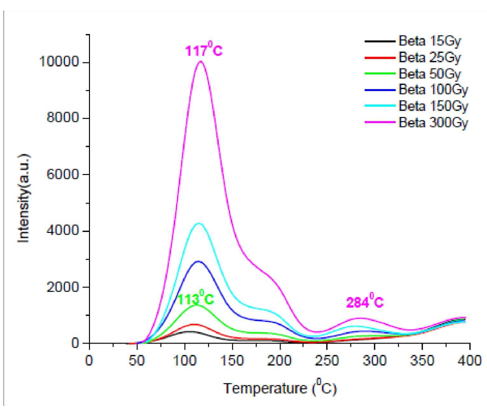

Fig.-3: TL glow curves of CaF_2 exposed to various amount of Sr-90 beta doses.

Table-2

Sr. No.	Beta dose (Gy)	Peak Temp. (^{0}C)	Peak Intensity (A.U.)
1.	15	105	428
2.	25	111	682
3.	50	113	1371
4.	100	115, 287	2924, 444
5.	150	115, 285	4221, 623
6.	250	150, 284	10037, 903

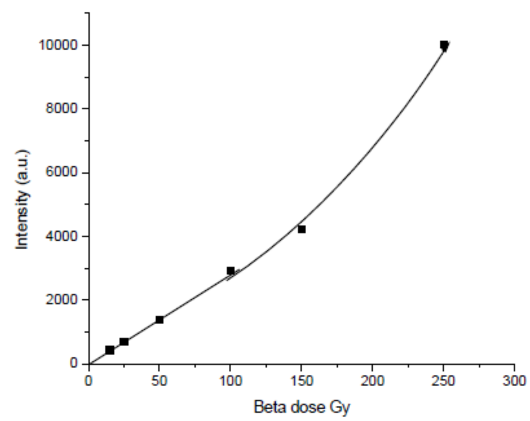

Fig.-4: Dose response curve of CaF_2

For the CaF, sample, the glow curve exhibits two primary peaks at ~105/ °C and 284/ °C, along with a broad hump around 190/ °C. Both the intensity of the hump and the low-temperature peak increase with increasing beta dose. Additionally, the low-temperature peak shows a shift from 105/ °C to 117/ °C as the dose increases. Notably, the sample undergoes a visible color change to cyan when the applied beta dose exceeds 100/ Gy, suggesting possible dose-induced changes in the defect centers responsible for TL emission. The TL emission characteristics of the 117/ °C peak may correlate with this color change. The corresponding TL peak temperatures and intensities for various doses are presented in Table 2.

These observations indicate that both dolomite and CaF_2 exhibit dose-dependent TL behavior, with clear shifts in peak temperature and significant increases in intensity. Such characteristics highlight their potential suitability for dosimetric applications across a range of beta radiation doses.

3.2 Thermoluminescence Decay

The thermoluminescence (TL) decay behavior of dolomite and calcium fluoride (CaF_2) samples was investigated to evaluate the stability of the stored TL signal with storage time after beta irradiation.

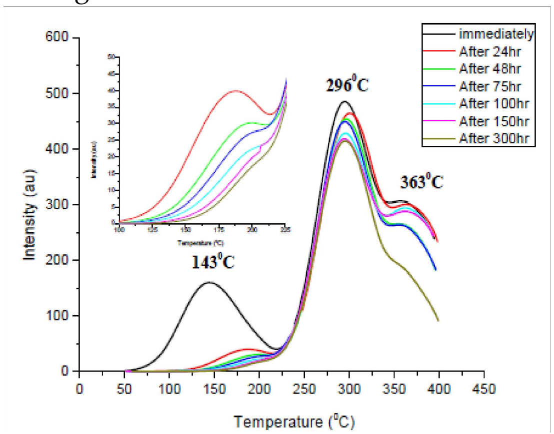

Fig.-5: TL glow curve of dolomite post exposure of beta radiation after different time

The TL glow curves recorded immediately after irradiation exhibit two distinct peaks at approximately 143 °C and 296 °C.

With increasing storage time, a significant reduction in the intensity of the low-temperature peak (~143 °C) is observed. After 48 h of storage at room temperature, this peak becomes poorly resolved and eventually disappears, indicating rapid fading of shallow traps associated with this peak. The gradual decrease in intensity of the low-temperature peak with storage time suggests thermal instability of the corresponding trapping centers.

Table-3

Sr. No.	Time of Storage (h)	Peak Temp. (^{0}C)	Peak Intensity (A.U.)
1.	Immediately	143, 296	160, 485
2.	24	296	454
3.	48	297	449
4.	75	296	435
5.	100	296	425
6.	150	297	420
7.	300	296	419

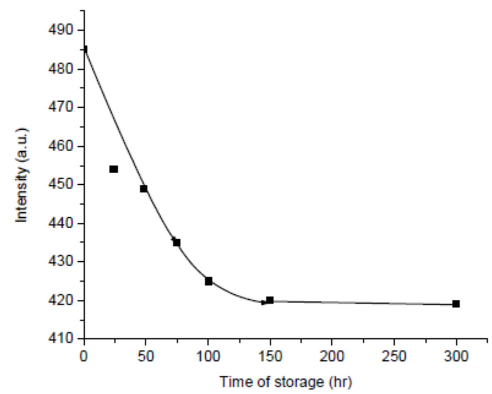

Fig.-6: TL decay curve for dolomite

In contrast, the high-temperature peak at ~296 °C shows comparatively better stability. Although its intensity decreases with increasing storage time, the peak remains clearly observable throughout the storage duration. Analysis of the TL decay curve reveals that the intensity of the 296 °C peak decreases progressively and is reduced by approximately 13 % of its initial intensity after prolonged storage. This relatively small reduction indicates that the deeper traps responsible for the high-temperature peak possess greater thermal stability and are less susceptible to fading effects.

The TL peak temperatures and corresponding intensities measured at different storage intervals following beta irradiation are presented in Table-3. The observed fading behavior highlights the importance of trap depth in determining signal stability and suggests that the high-temperature peak is more suitable for dosimetric applications requiring longer storage times between irradiation and readout.

Overall, the TL decay characteristics of dolomite demonstrate that while low-temperature peaks are prone to rapid fading, high-temperature peaks exhibit sufficient stability, making these materials promising candidates for thermoluminescence dosimetry.

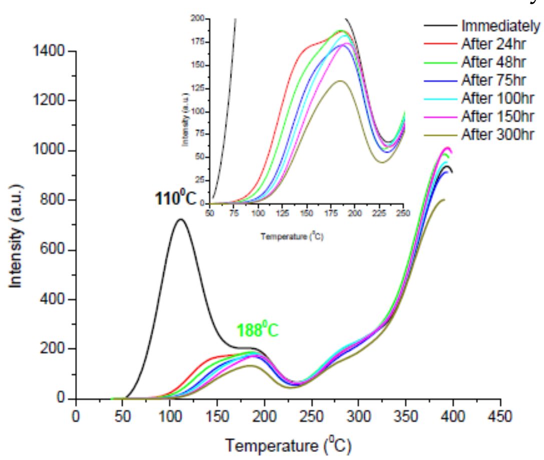

Fig.-7: TL glow curve of CaF_2 post exposure of beta radiation after different time

Table-4

Sr. No.	Time of Storage (h)	Peak Temp. (^{0}C)	Peak Intensity (A.U.)
1.	Immediately	110, 188	723, 193
2.	24	188	187
3.	48	188	184
4.	75	188	180
5.	100	188	170
6.	150	188	160
7.	300	188	140

The thermoluminescence glow curve of the natural calcium fluorite sample exhibits a prominent low-temperature peak at approximately 110 °C, along with a broad hump centered around 188 °C. These features indicate the presence of shallow and intermediate trapping levels within the crystal lattice.

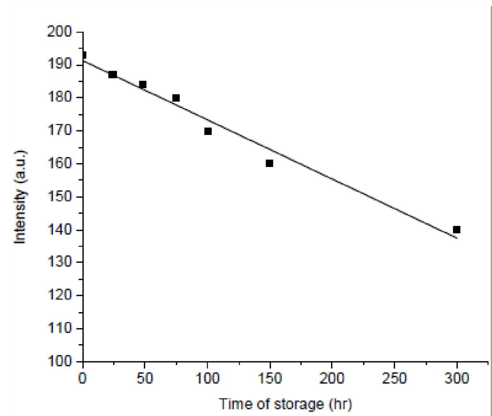

Fig.-8: TL decay curve for CaF_2

A pronounced fading effect is observed in the low-temperature peak with increasing storage time after beta irradiation. The intensity of the 110 °C peak decreases rapidly and is reduced by nearly 75 % of its initial value after 24 h of storage at room temperature. This rapid loss of signal suggests that the traps responsible for this peak are thermally unstable and undergo significant thermal release even at ambient conditions.

Top of FormFollowing 24 h of storage, the reduction in intensity of the 110 °C peak leads to the transformation of the hump around 188 °C into a broad, merged peak. This behavior can be attributed to the overlap of glow peaks and the depopulation of shallow traps, resulting in redistribution of charge carriers among deeper trapping levels.

The variation of TL peak temperature and peak intensity with storage time after beta irradiation is presented in Table-4. The observed fading characteristics demonstrate that the low-temperature TL peak in natural dolomite is highly susceptible to signal loss, limiting its usefulness for long-term dosimetric applications. However, the presence of broader, relatively stable features at higher temperatures may still be of interest for short-term or immediate readout dosimetry.

4. Conclusion

The thermoluminescence dosimetric behavior of natural dolomite and CaF, minerals was systematically studied through TL growth and decay analyses. Both materials exhibited well-defined glow peaks with increasing intensity and temperature shifts as a function of beta dose. Significant fading was observed for low-temperature peaks, whereas high-temperature peaks demonstrated improved thermal stability. These findings suggest that dolomite and CaF, , particularly their stable high-temperature TL peaks, are potential candidates for short- and medium-term radiation dosimetry.

5. Future Scope of Work

Future investigations may focus on kinetic parameter evaluation, TL emission spectroscopy, and dopant modification to enhance sensitivity and stability. Long-term fading studies and gamma irradiation response could further establish the dosimetric applicability of dolomite and CaF,.

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