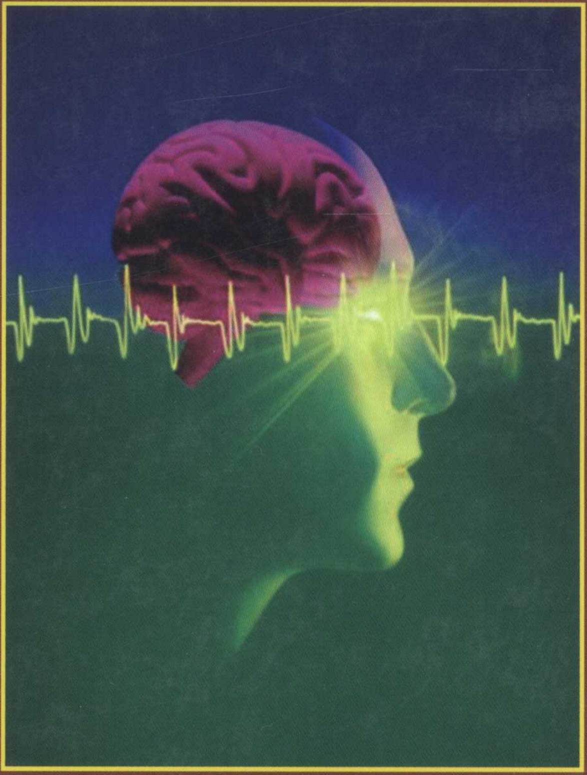

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LECTURES
EXERCISES

BIO MEDICAL PHYSICS

Chisinau • 2014

MINISTRY OF HEALTH OF REPUBLIC OF MOLDOVA
UNIVERSITY OF MEDICINE AND PHARMACY
"NICOLAE TESTEMITANU"
Department of Human Physiology and Biophysics

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PREFACE

Biophysics is the science that was developed through the interplay of two sciences, so-called "classical" – physics and biology –, from which comes the name.

Until now, the complete definition of biophysics has not reached to a consensus. So, we selected some definitions, which in our opinion, should be the most significant ones.

Professor W. Beier from Leipzig Medical University defines biophysics as "the science that deals with the physical analysis of functional structures and biological behavior". Professor V. Vasilescu from Bucharest University of Medicine and Pharmacy: "Biophysics is the science that studies physical phenomena from biological systems in light and with the aid of theories and physico-mathematical techniques". Professor N. Tarusov from Moscow "M. V. Lomonosov" University: "Biophysics consists of studying the phenomena and physico-chemical molecular structure underlying the primary mechanisms of biological processes".

The analysis of the proposed definitions allows us to conclude: **"Biophysics is the science dealing with physical analysis of the primary mechanisms which refers to composition, structure, existence and development of living matter"**. But biophysics does not reduce the problems of biology and medicine to simple physico-chemical ones. Life, distinguishing qualitatively from physico-chemical phenomena, follows specific laws that don't have any reason in lower areas of complexity.

From the wide variety of topics that biophysics has, medical biophysics requires the selection of problems connected with major directions of development of medicine.

This course is for 1st year students (departments: Medicine, Public Health, Stomatology and Pharmacy), being useful to everyone who wishes to make a clear picture of the essence of life processes.

The topics of the course were selected after analysis of traditional programs of similar courses from different universities, both from East and West. We wanted the structure and contents to be closer to what is done in medical institutes from France. Obviously, it was taken into consideration the correlation with the programs of disciplines, which students in medicine go through each year of study.

The content of course was influenced by the fact that, traditionally, some research methods, as for example, the spectral analysis, polarimetry and others, are fully studying in laboratory. On the other hand, having a short-term of teaching this subject, we simplified, as far as possible, the usage of mathematical apparatus.

Every theme of the course is followed by well-integrated exercises with its contents and with its direction vector of professional training.

We should recognize that the topics of medical biophysics courses, proposed by different authors, and of course of present course, have a subjective characteristic because it reflects, measurably, the inclinations and concerns of those who have made this choice.

Making available to students this manual, we hope that it attends not only a source of knowledge, but will be a catalyst in grafting of interest for fundamental sciences and a desire to work for a proficient application of achievements of physics and modern techniques in practice and medical research.

We consider that the illustrative component, which was thoroughly selected from the Internet with certain changes for correlation with the content theme, will generate an affective-positive effect, serving their students as impulse to familiarize themselves more deeply with the essence of physical phenomena involved harmoniously and economical in life processes.

Thank you to all the readers and we are waiting for critical suggestions and objections, oriented to correct and complete the content of the book, in a future edition.

Authors

1. THE STRUCTURE OF THE MATTER. BOHR'S ATOM

1.1. GENERAL NOTIONS

Whatever state of aggregation will be-solid, liquid or gas, the matter represents the substance which is the result of the association of identical molecules for pure or different bodies, for compound ones. Molecules are obtained by connection of several identical or different atoms. The atom is the elementary particle of any chemical element.

All chemical elements, identified till today, are assigned, according to electronically structure of the atoms and to their chemical properties, in the periodic table of **Mendeleev's**.

The atom of each element is perfectly defined. It is composed of a nucleus with a positive electrical capacity, around which are revolving the electrons with negative electrical capacity.

The nucleus is compound of **nucleons**: **protons** and **neutrons**.

Two numbers characterize any atom **X**:

– **A:** it represents the total number of nucleons from the nucleus; it is appointed **number of mass**, because, it reflects, practically, the hole mass of the atom;

– **Z**: it homologizes the number of protons from the nucleus; it's appointed **atomic number**, because, it is also the serial number in the periodic classification.

The difference **N = A – Z** homologizes the number of neutrons from the nucleus.

By convention, it was adopted the fallowing inscription:

$^{A}_{Z}\mathbf{X}$, where **X** represents the symbol of the atom.

For example: $^{4}_{2}\mathbf{He}$; $^{12}_{8}\mathbf{C}$; $^{235}_{92}\mathbf{U}$ etc.

Beacouse electrically the atom is neutral, the number of the electrons revolving around the nucleus dovetails with the number of the protons from the nucleus.

There are specific units of measure in atomical physics.

- **The atomic mass unit (a.m.u.).**

It consists of 1/12 from the carbon atom mass 12. When knowing the mass of the mole of carbon $^{12}_{8}$C is equal to 12 g, a.m.u. may be expressed in grams:

$$1\,a.m.u.=\frac{1}{12}\cdot\frac{12}{N_A}=\frac{1}{N_A}=\frac{1}{6.023\cdot 10^{23}}=1.66\cdot 10^{-24.}\,g.$$

Where: N_A – **Avogadro's** number.

- **The unit of electron-volt energy (eV)**

This unit corresponds to the kinetic energy of an electron, accelerated by the potential difference equal to 1 volt:

$$\mathbf{1eV = 1.6 \cdot 10^{-19} C \cdot 1V = 1.6 \cdot 10^{-19} J}.$$

The equivalence between the mass and the energy, expressed in the **Einstein's** relation: **$E = mc^2$**, allows the convertion of the mass units into energy units and vice versa. For exemple:

1 a.m.u. corresponds to the energy of 931.5 MeV.

Elementary particles in the atoms structure

The proton

According to the general denotation $^A_Z X$, the proton is represented $^1_1 p$; being the nucleus of hydrogen atom, it can be, also, represented by the symbol $^1_1 H^+$. By value, its positive charge it is that of the electron: **$1.6 \cdot 10^{-19}$ C.**

Its mass has the value: **$m_p = 1.007596$ a.m.u.** It is stable also outside the nucleus.

The neutron

It has the symbol $^1_0$n; it has no electric charge, and has the mass a little bigger than that of the proton: **$m_n = 1.008986$ a.m.u.**

Outside the nucleus, the neutron is radioactive and it disintegrates in the proton, the electron and the antineutrino, as the equation:

$$^1_0 n \rightarrow {}^1_1 p + {}^0_{-1} e + {}^0_0 \nu.$$

The electron

It has the mass (m) equal to **$9.1 \cdot 10^{-31}$ kg** and the charge (e) to **$-1.6 \cdot 10^{-19}$ C**.

The distribution of protons and neutrons into the nucleus allows the definition of the particular groups of mentioned elements.

The isotopes

There are elements which atoms have the same serial number **(Z)** and the same chemical properties, having a different number of mass **(A)**, on the strength of the different number of neutrons in nucleus.

They have different physical properties; particularly, some of them may be radioactive, being called **radioisotopes.**

For exemple, the hydrogen has 3 isotopes:

$^1_1 H$ – hidrogen; $^2_1 H$ – deuterium; $^3_1 H$ – tritium (radioactive).

The isobars

There are elements that represent a mathematic curiosity. Their atoms have the same mass number, but a different atomic number, with both different properties: physical and chemical.

For exemple:

$$^{14}_{6}C \text{ and } ^{14}_{7}N.$$

Izotons

There are elements which atoms in the nucleus have the same number of neutrons, but the chemical and physical properties- net different. For exemple:

$$^{14}_{6}C,\ ^{15}_{7}N \text{ and } ^{16}_{8}O;\ \mathbf{N=A-Z=8} \text{ – the same for all izotons.}$$

Isomers

There are two elements absolutly identical. Only one difference of energie allows to distinguish them. One of them it is at a metastable energetical level (**n**) superior to the another. For exemple:

$$^{80m}_{35}Br \text{ and } ^{80}_{35}Br;\ ^{99m}_{43}Tc \text{ and } ^{80}_{43}Tc.$$

According to the previously mentioned aspects, an extensive caracteristics of the hole chemical elements is exposed in the periodic nuclear tabel (*fig. 1.1*).

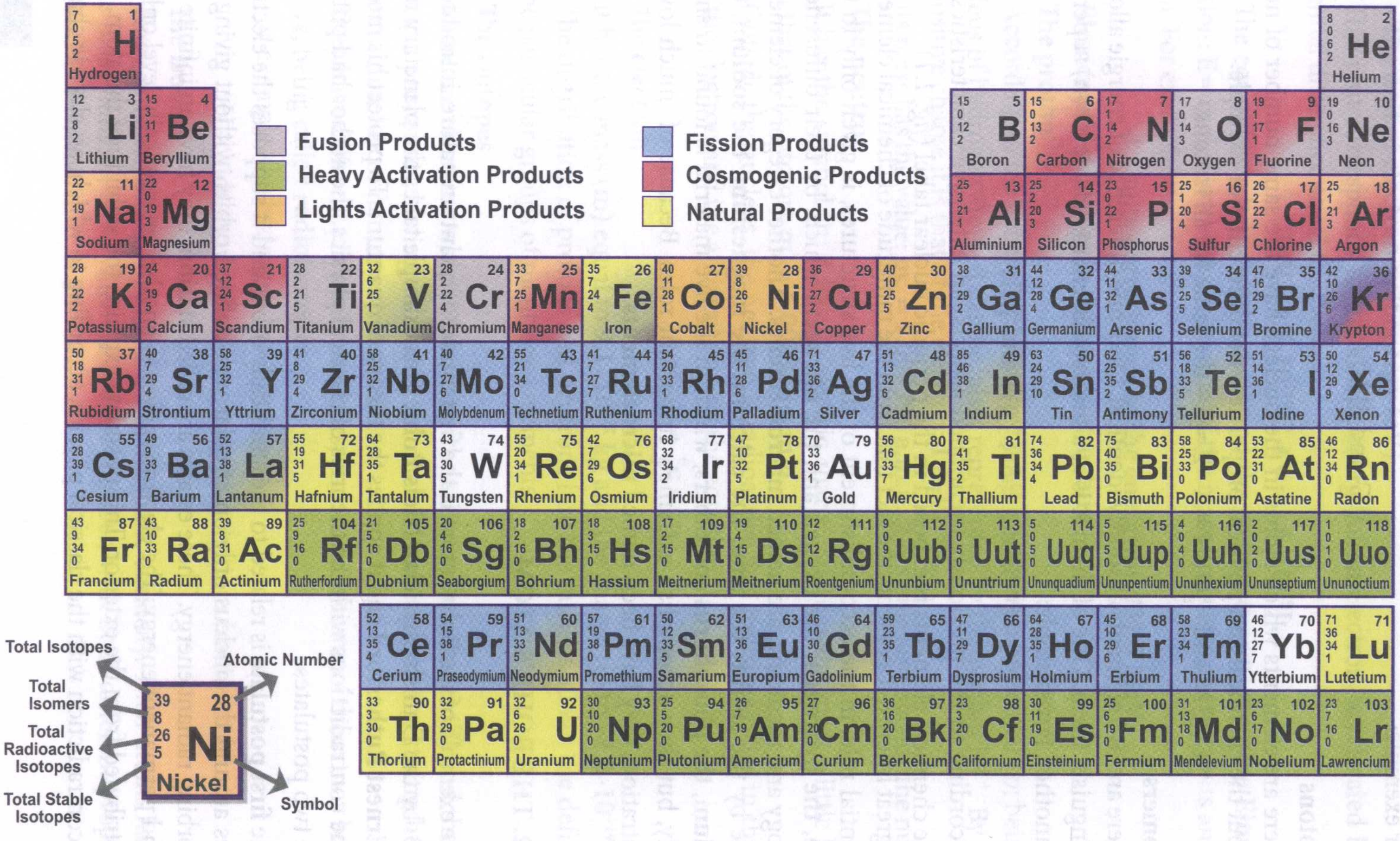
A great interest it presents the fact that from the multitude of chemical elements the esential role, in the accomplishment of the live structures, is given only to **the carbon**, **the hydrogen**, **the oxygen** and **the nitrogen**, which, by their diffuseness, in biology and medicine, are named **macroelements**. Another category of elements is made by **the sulfur**, **the iron**, **the calcium**, **the magnesium**, **the sodium**, **the potassium**, **the phosphorus** and others, which in the live material is found in small quantity, but they accomplish extremely important roles. Because of much lower concentration, they have been named **microelements**.

1.2. THE BOHR'S ATOM

The exemple of the Bohr's atom is the first model of quantum nature, introduced in 1913 by the Danish physician **Niels Bohr**. This example takes the planetary model of **Ernest Rutherford** and it applies the theory of quantum. To protect his model from the contradictions with the theories of classical physics, the author had putted to base two postulates.

The first postulate is related to the atomically orbits and it supposes the electron revolves around the nucleus only on circular well-defined orbits, without giving off or absorbing radiant energy. These estates are called **station**, they have an infinite life time and permanent energy. The electron moves to another energetically level only if it is distributed from the exterior. This postulate explains the stability of the atom, but it is in contradiction with the classical physics, according to which, an electric charge

Nuclear Periodic Table



Atomic Number	Symbol	Name	Total Isotopes	Total Isomers	Total Radioactive Isotopes	Total Stable Isotopes
1	H	Hydrogen	7	0	5	2
2	He	Helium	8	0	6	2
3	Li	Lithium	12	2	8	2
4	Be	Beryllium	15	3	11	1
5	B	Boron	15	0	12	2
6	C	Carbon	15	0	13	2
7	N	Nitrogen	17	1	14	2
8	O	Oxygen	17	0	14	3
9	F	Fluorine	19	1	17	1
10	Ne	Neon	19	0	16	3
11	Na	Sodium	22	2	19	1
12	Mg	Magnesium	22	0	19	3
13	Al	Aluminium	25	3	21	1
14	Si	Silicon	25	2	20	3
15	P	Phosphorus	23	0	22	1
16	S	Sulfur	25	1	20	4
17	Cl	Chlorine	26	2	22	2
18	Ar	Argon	25	1	21	3
19	K	Potassium	28	4	22	2
20	Ca	Calcium	24	0	19	5
21	Sc	Scandium	37	12	24	1
22	Ti	Titanium	28	2	21	5
23	V	Vanadium	32	6	25	1
24	Cr	Chromium	28	2	22	4
25	Mn	Manganese	33	7	25	1
26	Fe	Iron	35	7	24	4
27	Co	Cobalt	40	11	28	1
28	Ni	Nickel	39	8	26	5
29	Cu	Copper	36	7	27	2
30	Zn	Zinc	40	10	25	5
31	Ga	Gallium	38	7	29	2
32	Ge	Germanium	44	12	28	4
33	As	Arsenic	44	11	32	1
34	Se	Selenium	39	9	25	5
35	Br	Bromine	47	16	29	2
36	Kr	Krypton	42	10	26	6
37	Rb	Rubidium	50	18	31	1
38	Sr	Strontium	40	7	29	4
39	Y	Yttrium	58	25	32	1
40	Zr	Zirconium	41	8	29	4
41	Nb	Niobium	58	25	32	1
42	Mo	Molybdenum	40	7	27	6
43	Tc	Technetium	55	21	34	0
44	Ru	Ruthenium	41	7	27	7
45	Rh	Rhodium	54	20	33	1
46	Pd	Palladium	45	11	28	6
47	Ag	Silver	71	33	36	2
48	Cd	Cadmium	51	13	32	6
49	In	Indium	85	46	38	1
50	Sn	Tin	63	24	29	10
51	Sb	Antimony	62	25	35	2
52	Te	Tellurium	56	18	33	5
53	I	Iodine	51	14	36	1
54	Xe	Xenon	50	12	29	9
55	Cs	Cesium	68	28	39	1
56	Ba	Barium	49	9	33	7
57	La	Lantanum	52	13	38	1
72	Hf	Hafnium	55	19	31	5
73	Ta	Tantalum	64	28	35	1
74	W	Tungsten	43	8	30	5
75	Re	Rhenium	55	20	34	1
76	Os	Osmium	42	7	29	6
77	Ir	Iridium	68	32	34	2
78	Pt	Platinum	47	10	32	5
79	Au	Gold	70	33	36	1
80	Hg	Mercury	56	16	33	7
81	Tl	Thallium	78	41	35	2
82	Pb	Lead	74	36	34	4
83	Bi	Bismuth	75	40	35	0
84	Po	Polonium	58	25	33	0
85	At	Astatine	53	22	31	0
86	Rn	Radon	46	12	34	0
87	Fr	Francium	43	9	34	0
88	Ra	Radium	43	10	33	0
89	Ac	Actinium	39	8	31	0
104	Rf	Rutherfordium	25	9	16	0
105	Db	Dubnium	21	5	16	0
106	Sg	Seaborgium	20	4	16	0
107	Bh	Bohrium	18	2	16	0
108	Hs	Hassium	18	3	15	0
109	Mt	Meitnerium	17	2	15	0
110	Ds	Meitnerium	19	4	15	0
111	Rg	Roentgenium	12	0	12	0
112	Uub	Ununbium	9	0	9	0
113	Uut	Ununtrium	5	0	5	0
114	Uuq	Ununquadium	5	0	5	0
115	Uup	Ununpentium	5	0	5	0
116	Uuh	Ununhexium	4	0	4	0
117	Uus	Ununseptium	2	0	2	0
118	Uuo	Ununoctium	1	0	1	0
58	Ce	Cerium	52	13	35	4
59	Pr	Praseodymium	54	15	38	1
60	Nd	Neodymium	51	13	33	5
61	Pm	Promethium	57	19	38	0
62	Sm	Samarium	50	12	33	5
63	Eu	Europium	51	13	36	2
64	Gd	Gadolinium	46	10	30	6
65	Tb	Terbium	59	23	35	1
66	Dy	Dysprosium	47	11	29	7
67	Ho	Holmium	64	28	35	1
68	Er	Erbium	45	10	29	6
69	Tm	Thulium	58	23	34	1
70	Yb	Ytterbium	46	12	27	7
71	Lu	Lutetium	71	36	34	1
90	Th	Thorium	33	3	30	0
91	Pa	Protactinium	32	3	29	0
92	U	Uranium	32	6	26	0
93	Np	Neptunium	30	10	20	0
94	Pu	Plutonium	25	5	20	0
95	Am	Americium	26	7	19	0
96	Cm	Curium	27	7	20	0
97	Bk	Berkelium	36	16	20	0
98	Cf	Californium	23	3	20	0
99	Es	Einsteinium	30	11	19	0
100	Fm	Fermium	25	6	19	0
101	Md	Mendelevium	31	13	18	0
102	No	Nobelium	23	6	17	0
103	Lr	Lawrencium	23	7	16	0

Fig. 1.1.

in accelerated movement, as the movement of the electron, issues electromagnetic radiation. This is the reason of the decrease of the system energy, and the circular orbit of the electron would have the radius smaller and smaller, until it would „fall" on the nucleus. But experimentally is found the atom is stable and has particular estates in which its energy keeps itself constant.

The second postulate states the fact that the atom gives off or absorbs the electromagnetic radiation only in transition from one stationary estate to another. The energy which it receives or it offers is equal to the difference of the two levels between which the transition is made.

The Bohr's atomic example is applicable to hydrogen atoms, as to the hydrogenous ions He^+, Li^{2+}, Be^{3+} etc., (namely, to the ions that have only one electron).

The choice of the circular orbits admitted for the movement of the electron has at the base **Louis Victor de Broglie**, hypothesis, to which, every particle in movement has also, undulating properties. The wavelength **λ**, associated to one particle with the mass **m** and the movement speed **v**, is found from the relation:

$$\lambda = \frac{h}{\mathrm{mv}} \quad (1).$$

Where: $\boldsymbol{h} = \mathbf{6.62 \cdot 10^{-34}}$ **J·s** is **Planck's** constant.

The orbit is accepted if it has a whole number of wavelength **λ**. For a circular orbit in the situation of Bohr's atom, we obtain:

$$2\pi\mathrm{r} = n\lambda \quad (2).$$

Where: $\boldsymbol{n}$ – a full positive number; λ – the associated wavelength; **r** – the orbit radius.

Substituting the expression for λ from (1) in (2), we get:

$$2\pi\mathrm{r} = n\frac{h}{\mathrm{mv}} \Rightarrow \mathrm{mvr} = n\frac{h}{2\pi} \quad (3).$$

The term **L = mvr** is **the orbital kinetic moment of the electron**. Eventually (3) is the quantification relation of the kinetic moment.

The radii of accepted orbits are also quantification according to the relation:

$$r = \frac{n^2 h^2 \varepsilon_0}{\pi e^2 m} \quad (4).$$

Where: $\boldsymbol{e}$ – the electron charge; $\boldsymbol{\varepsilon_0}$ – the electric permittivity of the vacuum.

There are also quantification the energies on different levels:

$$W_n = -\frac{me^4}{8\varepsilon^2 h^2} \cdot \frac{1}{n^2} \quad (5).$$

Each stationary estate corresponds to a special energetical estate of the atom.

The atom passes from one stationary estate to another, with superior energy, only if it is given one quantum of appropriate energy with the difference between those two levels. When it comes back to the inferior level is issued the same radiation as at the absorption. This fact explained the discontinuous nature of the mass and the energy at the microscopic level.

When the atoms absorb the energy, the electrons make transitions from the nucleus, but when they issue the energy in the form of photons, the transitions of the electrons happen to the nucleus (*fig. 1.2*). The energy of the issued photons is determined from the relation:

$$h\nu = Wni - Wnf \quad (6).$$

Where: ***ni*** – the initial level; ***nf*** – the final level.

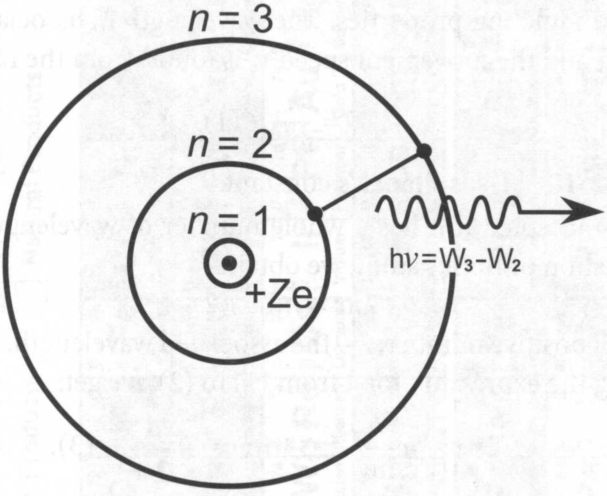


Fig. 1.2.

In the hydrogen spectrum had been detected more series of lines (*fig. 1.3*), which had confirmed Bohr's model.

The Lyman's series can be formed after the transitions between any extern level and the first one (**n = 1**).

The Balmer's series appears after the transitions between any other extern level and the second one (**n = 2**), and **the Paschen's series** – after the transitions between any other extern level and the third one (**n = 3**).

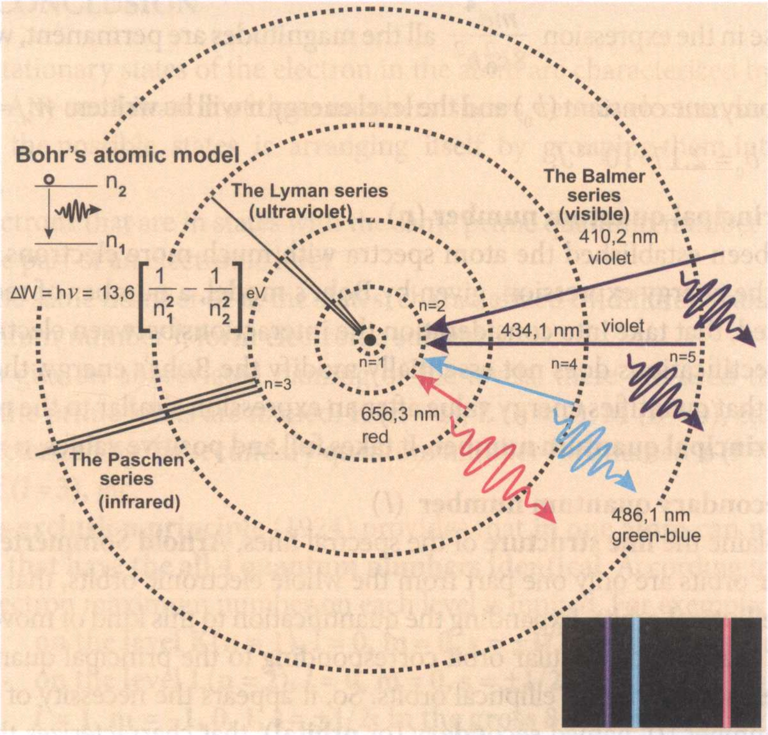


Fig. 1.3.

1.3. QUANTUM NUMBERS

Bohr's model is a simple model of quantized atom, considered a system maid from a nucleus and an electron that revolves around the nucleus on a **circular trajectory**. This model explains persuasively the emission and the absorption effects of the hydrogen atom.

When speaking about the atoms with more electrons, Bohr's model can not more explain the complex issue and absorption spectrums of these. In the atom with much more electrons, these interact not only with the nucleus, but, also, between them. Because of this, the energy expression of levels, concluded for the hydrogen atom and the hydrogenous ions, is not more correct. For the atoms with much more electrons, the energy quantification of levels is maid according to the relation (7), the periphery electron is attracted by Z positive charges:

$$W_n = -\frac{mZ^2e^4}{8\,\varepsilon_0^2\,h^2}\cdot\frac{1}{n^2} \quad (7).$$

Because in the expression $\frac{me^4}{8\varepsilon_0^2 h^2}$ all the magnitudes are permanent, we will note it through only one constant (b_0) and the level energy **n** will be written: $W_n = -b_0 \cdot \frac{Z^2}{n^2}$; calculated $b_0 = 2.17 \cdot 10^{-18} J$.

The principal quantum number (n)

It has been established the atom spectra with much more electrons can be explained if the energy expression, given by Bohr's model, a number of rectifications is introduced, that take into consideration the interactions between electrons. Because these rectifications does not essentially modify the Bohr's energy, the quantum number **n**, that quantifies energy value after an expression similar to the relation (7), is named **principal quantum number**. It takes full and positive values: **n = 1, 2, 3, ...**

The secondary quantum number (*l*)

To explaine the **fine structure** of the spectral lines, **Arnold Sommerfeld** believed the **circular** orbits are only one part from the whole electronic orbits, that have to include also **elliptical** orbits. Expanding the quantification to this kind of movement, was established that to each circular orbit corresponding to the principal quantum number **n** it belongs a series of elliptical orbits. So, it appears the necessity of the second quantum number (***l***), named **secondary** (or **orbital**), that charracterizes the kinetical moment of the electron corresponding to different elliptical orbits of the same level.

For the same **n**, ***l*** can get the amounts **0, 1, 2, 3... n–1**; so, to one level **n** of energy it corresponds **n** elliptical orbits.

The magnetic quantum number

When to explain the **giromagnetic** phenomena (the relation between the magnetic and kinetic moment) of the atoms, it was necessary to take into consideration the fact that electronical orbits are not in the same plan, but have a spatial allocation, so it brought to the introduction of the third **quantum number (m)**, named **magnetic**.

For a certain number of ***l***, the magnetic quantum number can get **2*l* + 1** numbers, so it fallows:

$$-l, \ldots, (-1), 0, 1, \ldots l.$$

The spine quantum number (s)

Different spectra particularities of the **alcaline metals** of the leaded (1925) to the hypothesis that the electron has also, a rotation round one of its own axes, movement named **spine.**

The spine quantum number (**s**), takes for the electron the values ± ½, according to the rotation direction round its own axes.

1.4. CONCLUSION

The stationary states of the electron in the atom are characterized by 4 quantum numbers. Any combination of these numbers shows certain electron state. The multitude of the possible states is arranging itself by grouping them into **levels** and **sublevels.**

All elctrons that are in states with the same prime quantum number are considerated to be part of an electronic level.

For the same number of **n**, the states, characterized by different values of secondary quantum number ***l***, form electronic sublevels.

As to give an abbreviated naming of the levels, there are used the following symbols: the prime levels are named: **K (n = 1)**; **L (n = 2)**; **M (n = 3)**, etc.; the sublevels characterized by the secundary quantum number l are named: **s** ($l = 0$); **p** ($l = 1$); **d** ($l = 2$); **f** ($l = 3$), etc.

Paul's exclusion principle (1924) provides that **in one atom can not exist two electrons that have the all 4 quantum numbers identical**. According to this principle, the electron maximum number on each level is limited. For exemple:

on the level K(n = 1), $l = 0$, m = 0, s = ±1/2; in the gross 2 electrons;
on the level L(n = 2), $l = 0$, m = 0, s = ±1/2;
$l = 1$, m = –1, 0, 1, s = ±1/2; in the gross 8 electrons;
on the level M(n = 3), $l = 0$, m = 0, s = ±1/2;
$l = 1$, m = –1, 0, 1, s = ±1/2;
$l = 2$, m = –2, –1, 0, 1, 2, s = ± 1/2; in the gross 18 electrons, etc.

From the ones related above, it comes out the general rule: on the **n** level can be **$2 \cdot n^2$** maximum.

EXERCISES

1. Determine the order number (Z) of an element in the periodic system, knowing the charge of the atomical nucleus is equal to $8 \cdot 10^{-18}$ C.

Answer: Z = 5.

2. Calculate in Å the wavelength associated to an electron that has the speed of $3 \cdot 10^7$ m·s^{-1}; ($h = 6.6 \cdot 10^{-34}$ J·s; $m_e = 9.1 \cdot 10^{-31}$ kg).

Answer: λ = 0.24 Å (radius X).

3. Calculate in eV the whole energy of the electron in the hydrogen atom on the first energetical level.

Answer: –13.6 eV.

4. Calculate the energies of the first two energetic levels in keV and the wavelength in Å of radiation, which occurs after the transition of an electron from level 2 on level 1 (Z = 74).

Answer: $W_1 = -74.27$ keV; $W_2 = -18.57$ keV; $\lambda = 0.22$ Å.

5. Calculate the frequency ν in s^{-1} and wavelength λ in Å of the radiation emitted by hydrogen atoms when the electron from the level M returns to the level K.

Answer: $\nu = 0.29 \cdot 10^{16}\ s^{-1}$; $\lambda = 1034$ Å (ultraviolet).

6. An electron with the initial velocity $v_0 = 0$ was accelerated by a potential difference $5 \cdot 10^4$ V. Which is the wavelength of the photon, the energy which would be equal to the kinetic energy of the electron?

Answer: $\lambda = 0.25$ Å.

7. In an oscilloscope tube, accelerating voltage is 20kV. Which is the wavelength associated to the electron in the finishing acceleration?

Answer: $\lambda = 8.86$ pm.

8. Which wavelength has the radiation emitted by the hydrogen atom at the electron transition on an inner orbit, releasing energy of 1.892 eV?

Answer: $\lambda = 656.3$ nm.

2. ELEMENTS OF MOLECULAR BIOPHYSICS

2.1. INTERATOMIC AND INTERMOLECULAR FORCES AND BONDS

Interatomic and intermolecular forces derive from the fact that the atom of all chemical elements containing electrical charges of opposite sings: positively charged nucleus due to protons and negatively charged electron shell. Therefore these electrostatic forces are likely based on **Coulomb's law:**

$$F = \frac{1}{4\pi\varepsilon_0} \cdot \frac{q_1 \cdot q_2}{\varepsilon r^2} \quad (1).$$

Where: $\boldsymbol{\varepsilon_0}$ – electric permittivity of vacuum; $\boldsymbol{\varepsilon}$ – relative permittivity of a dielectric medium in comparison with the reference medium (vacuum).

The expression $\frac{1}{4\pi\varepsilon_0}$, is a constant size, noted by K. Law, from the mathematical point of view, write:

$$F = K\cdot\frac{q_1\cdot q_2}{\varepsilon\cdot r^2}\quad (2);\quad \text{counted } K = 9\cdot 10^9.$$

It is obvious that the relative permittivity of vacuum is equal to 1, other dielectric medium possessing a higher relative permittivity than 1.

Electric field intensity $\vec{E}$, created by a point charge q at a distance r from this charge is defined by the relation $E = K\frac{q}{\varepsilon r^2}$, having as a unit newton · coulomb^{-1} (N · C^{-1}) or volt · meter^{-1}(V · m^{-1}).

The relation defines electrostatic potential created by an electric charge punctiform distance r is defined by the relation: $V = K\frac{q}{\varepsilon r}$, measuring unit is volt (V), from the name of Italian physician A. Volta (1745–1827).

When charges have the same sign, the interaction forces are rejecting, and when they have opposite signs – the attraction.

Stable configuration, called **molecule**, is formed when the forces of attraction between the constituent atoms are approximately equal to the forces of rejection and the potential energy of the system is minimal.

To elucidate the mechanism of molecular components of biological structures is necessary to know the different types of links between these forces and provide these links.

From interactions at the atomic-molecular level are mention the linkages: **Van der Waals; the hydrogen bond; ionic and hydrophobic.**

Van der Waals connection (bond)

This bond is relatively weak and can be made between atoms with full outer electronic shells or saturated chemical bonds between molecules.

Size of forces which provided that the period is characterized by a/V_0^2 from gas equation for one mole of gas:

$$(P + a/V_0^2)(V_0 - b) = RT\quad (3).$$

Where ***a*** and ***b*** are experimental constants, that reflect real gas properties in relation to perfect gases.

The term a/V_0^2 reflects pressure correction taking into account the attraction of molecules due to intermolecular forces.

Constant ***b*** is the volume correction, taking into account volume of the gas molecules if perfect gas is neglected.

Van der Waals forces do not change the structure and properties of molecules, just sitting in the formation of molecular complexes especially biological macromolecules.

Neutral molecules can have a symmetric spatial structure in terms of power (molecule of methane – **CH_4**) or asymmetric structure in which tasks are separated spatially opposite (water molecule – **H_2O**), the last is **permanent electric dipole.**

A dipole is a structure formed by two point electric charges of the same magnitude, but opposite sign, located at a distance (*l*) (*fig. 2.1*).

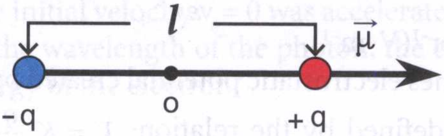


Fig. 2.1.

Degree of polarization of a molecule is measured by its electric moment ($\vec{\mu}$), named also **dipole moment**:

$$\vec{\mu} = q\,l \quad (4).$$

Conventional direction is significant from the negative charge to the positive, in extension of dipole axis. The unit of dipole moment in S.I. is coulomb·meter (C·m). Basically it uses a much smaller unit called *debye* (D):

$$1D = 1/3 \cdot 10^{-29}\ C \cdot m\,.$$

Between dipoles molecules are exercised electrostatic forces, named **Van der Waals.**

Under the action of an external electric charges and dipoles, molecules can become symmetrical structure. Such a dipole is called **dipole induced** (*fig. 2.2*).

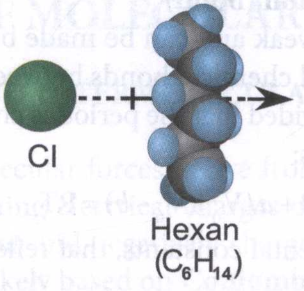


Fig. 2.2.

Neutral molecules are dipoles that may become due to the movement of electrons. Such dipoles have directly influenced only in a very short time, being called **electric dipole moment**. The dipole formation mechanism can be demonstrated by research interaction between two helium atoms. At a small distance between the atoms can be established between electron spin correlations that lead to the appearance of Van der Waals forces of attraction or rejection (*fig. 2.3*).

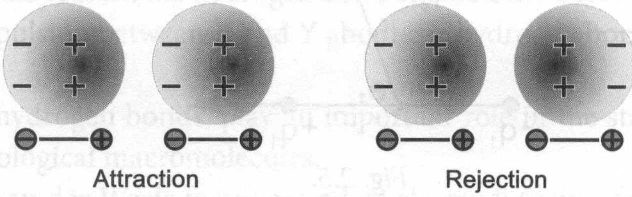


Fig. 2.3. Momentary dipoles.

Dipole–field interaction

Either a rigid or fixed electric dipole placed in an electric field with intensity $\vec{E}$ (*fig. 2.4*).

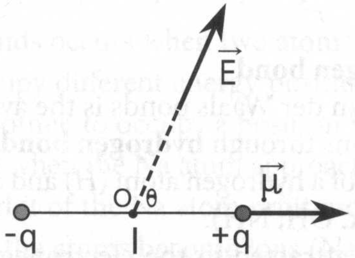


Fig. 2.4.

Potential energy of the dipole interaction with external field is determined from the relation:

$$W_p = -\vec{\mu} \cdot \vec{E} \quad (5).$$

Scalar product of these two vectors, by definition, depends on the cosine angle between their directions:

$$W_p = -\mu \cdot E \cdot \cos\Theta \quad (6).$$

Consequently, the potential energy of interaction dipole-field is minimal (stable configuration), when the dipoles of the field vectors are parallel and face the same direction. Configuration becomes unstable equilibrium (high potential energy), when these vectors are parallel and oriented in opposite directions.

Dipole–dipole interaction

To simplify the problem, we consider two rigid dipoles placed in the same plane (*fig. 2.5*).

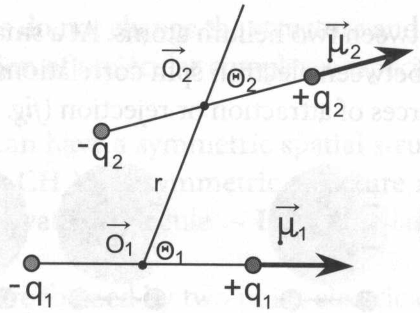


Fig. 2.5.

The relative position of two dipoles is determined by **r** (distance between the centers of dipoles) and angles θ_1 and θ_2 (between vector $\overrightarrow{O_1O_2}$ and dipole moments, respectively $\vec{\mu}_1$ and $\vec{\mu}_2$). Potential energy of interaction of these two dipoles in vacuum is determined from the relation:

$$W_p = -K\frac{\mu_1 \cdot \mu_2}{r^3} \cdot (2\cos\theta_1 \cos\theta_2 - \sin\theta_1 \sin\theta_2).$$

Links through hydrogen bond

A particular case of Van der Waals bonds is the average distance (about 3Å = $3 \cdot 10^{-10}$ m) is the connections through **hydrogen bond**.

A molecule consisting of a hydrogen atom (*H*) and an electronegative atom (*X*) shows a **covalent bond** (ex. OH, NH).

Hydrogen electron is attracted to the electronegative atom, leading to the formation of **permanent electric dipole**. If this dipole appears near another electronegative atom (*Y*), between XH and Y is realized now by **hydrogen bond** (*fig. 2.6*).

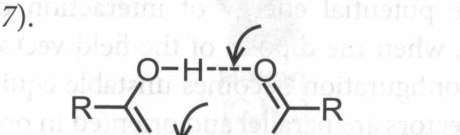


Fig. 2.6. Hydrogen bond with dipole-dipole interaction.

Scheme can serve as a parable of the bridge hydrogen bond occurs between carboxyl groups (*fig. 2.7*).

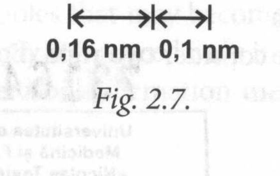


Fig. 2.7.

Hydrogen bond has the following features:

– as the nucleus of an atom of hydrogen (proton) is smaller than any atom, the atoms X and Y are so close, exciting another electronegative atom can no longer insert. For this reason, the hydrogen bond **can be only between two atoms**;

– due to repulsion between X and Y about the hydrogen bond is **always in a straight line**.

Linked by hydrogen bonds, play an important role in the stability of spatial structures of biological macromolecules.

Although Van der Waals forces are relatively small their effect becomes significant due to the large number of interactions of this kind. Ties based on these strengths allow breaking and training them quite easily, which condition the biological functions of biopolymers.

Ionic bonds

Formation of ionic bonds occurs when two atoms approach, in the structure of which the electrons occupy different energy positions, so that an electron can move from one atom to another, to occupy a position with lower energy.

An example is the case when the **Na** atom approaches an atom of **Cl**. The single electron in the outer orbit of the Na atom orbit moves into the external orbit of Cl atom. Consequently, the atoms become ions (Na^+ and Cl^-), between which electrostatic attractive forces appear keeping them connected to each other. Such a bond is called **ionic**. Ionic bond dissociates slightly in solution.

In fact, the formation of ions in solution is the most reliable evidence that the molecule in undissolved state contains an ionic bond. Crystals, in which this type of bond predominates, are called ionic crystals. Ionic bond is stronger than Van der Waals bonds, fact that explains the hardness of ionic crystals.

Hydrophobic bonds

Weak hydrophobic interactions are bonds existing between molecules or groups of molecules that are also non-polar. Although hydrophobic forces are rejecting, in the case when the number of molecules is large, interacting mutually, they lead to the formation of structures called **micelles**, in equilibrium state (*fig. 2.8*).

It is important that the bonds between the hydrophobic phospholipids molecules provide the structural organization of cell membranes.

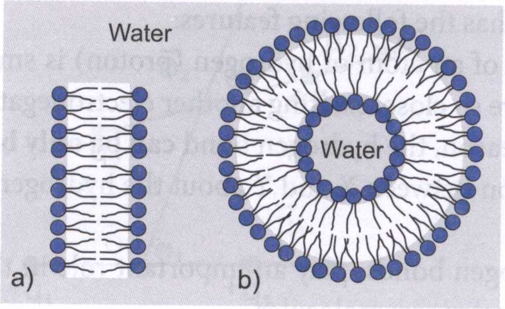


Fig. 2.8.

2.2. PHASE TRANSFORMATIONS. LIQUID CRYSTALS

When a substance passes from one aggregation to another, the process is named **phase transformation**. It occurs under certain conditions of temperature and pressure.

Triple point (TP) is where all three states of aggregation, in dynamic equilibrium, coexist and is characterized by a certain temperature and pressure specific for each substance. For example, water, TP is characterized by pressure (P) 609 Pa and temperature (T) 273.16 °K, for the carbon dioxide $5.16 \cdot 10^5$ Pa and 216.6 °K.

The critical point (CP) is the temperature over whose gas can't be liquefied, no matter how much the pressure increases.

With some common features, gases and liquids are called generic **fluids**. For example, water, CP corresponding to temperature (T) 647.16°K, air – 139.16°K, nitrogen – 126.16 °K.

In figure 2.9 is presented a general phase diagram.

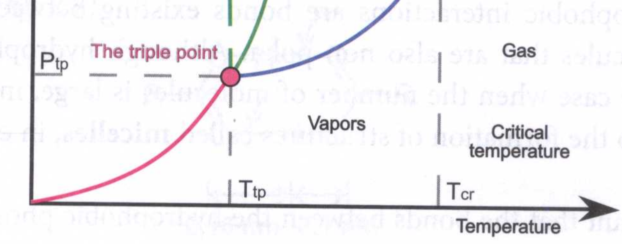


Fig. 2.9.

It is known that crystalline substances, in contrast to isotropic substances, have a specific temperature to change from a solid phase to a liquid on under atmospheric pressure (melting point).

Liquid crystals are organic substances in which the liquid-solid transition is not direct; instead they present an intermediate state with the same properties of a conventional liquid and solid crystal (anisotropic substance).

Intermediate stages are called: **mesophases**, **mesomorph** or **fluid condensed phase spontaneous anisotropy**.

Liquid crystals were discovered by Austrian botanist **F. Reinitzer** in 1888 and called **"liquid crystals"** by the German physicist **O. Lehman**.

Long though, physicists as well as chemists have not considered a stable thermodynamic state of matter, leading to a poor feasibility in practice. It was not until around 1930, suspecting the possibility of applications increased scientific concern mesomorphic materials scientists. During this period, the research conducted to reveal the structure and some properties of liquid crystals.

Today, there isn't a university or institute of basic and applied research in physics, which have no concerns about liquid crystals.

There are two major classes of liquid crystals: **liotrope** and **termotrope**.

Termotrope are substances that pass through the liquid-crystal phase by changing the temperature, and those that pass through the liquid crystals phase by changing their concentration in solution are called **liotrope**.

Termotropes are important from the point of view of fundamental research, as well as the technological applications and liotropes because of their role in biological systems and living tissue.

Termotrope liquid crystals are obtained by heating the solids and organic compounds that consist of long molecules or sometimes with disc-shaped molecules. Some termotrope materials under the heat suffer several mesophases in order to change from a solid state to liquid. Such materials are called **polymorphs**, and the phenomenon – **polymorphism**.

The classification of liquid termotrope crystals, was made by **G. Friedel** in 1922, essentially relies on the symmetry of their structure.

Depending on the spatial orientation of the molecules, liquid crystals termotrope are divided into three groups (*fig. 2.10*).

Nematic (gr. ***nima*** – thread) – substances in which the long axis of the molecules are aligned in a preferred direction, without appearing a layered structure (*fig. 2.10, a*).

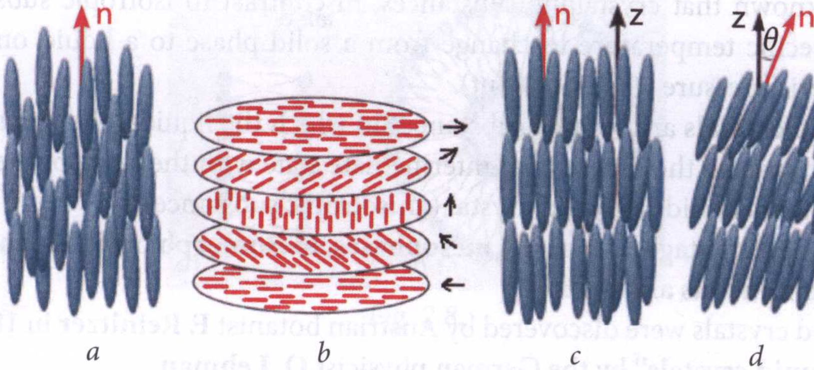


Fig. 2.10. Termotrope liquid crystals.

Observed under the microscope, these substances appear as thread – like strings attached. The direction of orientation of the molecule, so called **the optical axis of the crystal**, can be modified by the action of external factors, such as temperature, electric and magnetic fields, interactions with the walls of the enclosure in which they are. Nematic liquid crystals are almost 100% transparent, are devoid of optical activity and keep targeting molecules long time.

Cholesteric are the substances in which the long axes of the molecules are oriented parallel rotating continuously at an angle from one layer to another, forming a **layered structure** and **helical** (*fig. 2.10, b*). Due to this structure, they possess large number of properties.

Cholesteric liquid crystals are optically active (rotate the plane of polarization of light) are able to change color in the entire visible field, they depend on the temperature, which makes them useful **indicators of temperature in medicine**.

The temperature variation changes the cholesteric spiral step and consequently also the reflection of light and color, reflected from it, which is estimated to note that the crystal color change.

There are cholesterics with high sensitivity to temperature, which can detect temperature variation. For this purpose it is used encapsulated cholesteric polymer thin films. Applying these films on the body, color panel temperature distribution can be detected. In some cases, for the same purposes special ointments containing finely dispersed cholesteric liquid crystals are used on the skin. The method of visualization of temperature fields on the body is called **thermography** (*fig. 2.11*). Through this method, hidden inflammation of the skin centers and cancer formations can be detected.

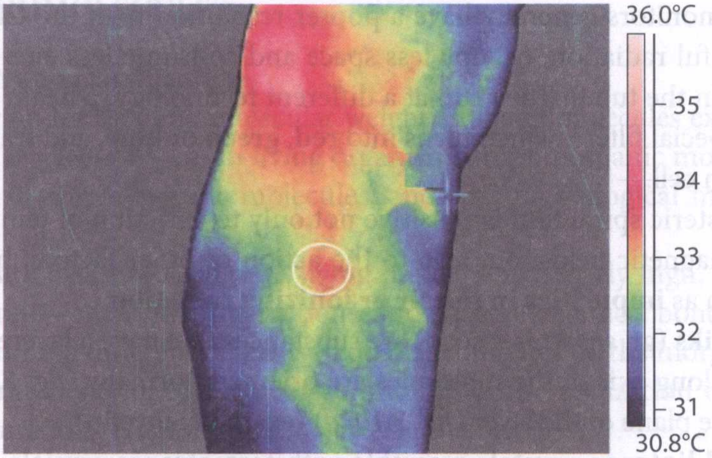


Fig. 2.11.

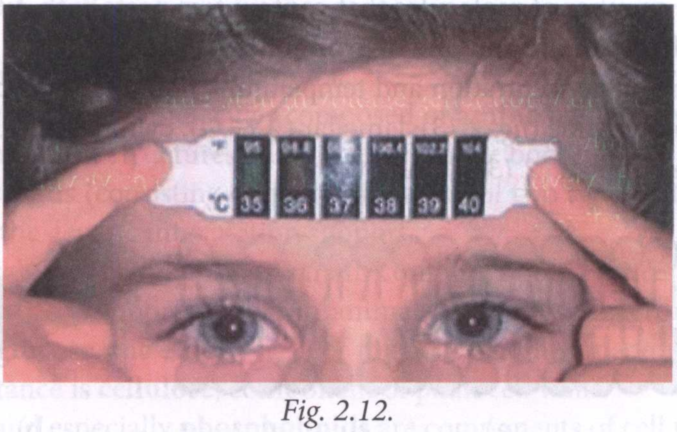


Fig. 2.12.

Films with cholesterol used as thermometers, are comfortable and free of danger (*fig. 2.12*).

Local heating of the „cholesteric" screen, obtained by transforming electrical impulses, promotes the use of colored TV, as well as processing and display system information (tonometers, clocks, monitors, etc).

Monitor output is a repository of car displays images and text. There are several types of monitors: Cathid ray tube (CRT), liquid crystal display (LCD), and plasma.

CRT monitors consist of a tube, which has at one end an electron gun and at the other end a screen with a luminescent coating. It operates under electron bombardment screen so it displays points of different colors.

LCD monitors generally have a poorer resolution than the CRT, but do not emit harmful radiation, occupy less space and consume less energy (about 5 W to 100 W in the tube). They run at a different technology: a beam of light passes through special filters, which turns into **red**, **green** or **blue**, and transmits electricity to each cell.

Cholesteric spiral film is sensitive not only to variation of temperature, electric and magnetic fields, but also to the action of other factors, harmful to humans, such as **impurities in the air or ionizing radiation.**

Smectiks (gr. ***smektos*** – soap) are substances that have a layered structure, in which the long axis of the molecules are oriented normally (*fig. 2.10, c*) or at an angle to the plane of the layer (*fig. 2.10, d*) (example: myelin).

Liquid liotrope crystals resemble with termotropes crystals by the fluidity and orientation of molecules, but differ in composition, having structures consisting of a large number of molecules that contain two parts with different characteristics: **one hydrophilic** and **one hydrophobic.**

Depending on concentration and temperature, liquid crystals can be of different structures: liotrope (*fig. 2.13*): lamellar (*2.13, a*), cylindrical (*2.13, b*), spherical (*2.13, c*), hexagonal (*2.13, d*) and others.

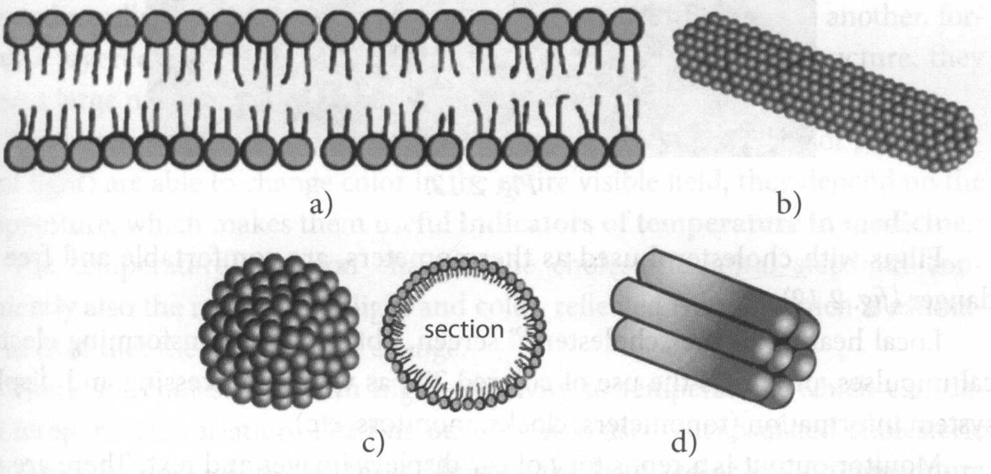


Fig. 2.13.

Lamellar structures present a very special property; the molecules are arranged in double layers, so that the hydrophobic ends inward without being in contact with water, and the hydrophilic heads – outward, being in contact with water (for example: cell membranes).

2.3. MACROMOLECULES

Definition and classification

As a criterion for defining molecular weight all macromolecules exceed 5 000 Daltons (D); $\mathbf{1D = 1.6 \cdot 10^{-27}\ kg}$. In living organisms from inorganic molecules are synthesized first simple organic molecules, and specific biological macromolecules.

The diversity macromolecules in living matter is extremely high, but nevertheless the atomic – molecular interactions, and types of chemical bonds in living matter, have nothing different in terms of quality compared to the inorganic. One of the main arguments in favor of these findings is the realization of artificial synthesis of numerous biomolecules.

Biological macromolecules (often called *biopolymers*) artificially synthesized have the same kinetic, electrical and optical properties as macromolecules synthesized in the cells.

Fundamental types of biopolymers:

1. **Proteins** are macromolecules composed of **amino acids**, they are involved in the most important structures and processes in the body.
2. **Nucleic acids** (consisting of **nucleotides**) are of two types: DNA and RNA, they make up the genetic information support.
3. **Polysaccharides: amylose**, the main component of plant cells, and **glycogen**, which plays an important role in animal cells. These macromolecules are used to store **glucose**, the „food" base of the cell. Another macromolecule with similar importance is **cellulose**, component of plant cell walls.
4. **Lipids** and especially **phospholipids** are components of cell membranes.

There are several criteria for the classification of biological macromolecules.

Depending on alternating monomers: any biopolymer consists of a series of rings, the basic units called **monomers**.

Macromolecular compounds called polymers are composed of groups of identical atoms. A polymer can be written as:

$$\mathbf{A - A - A - A - A} \text{ or } \mathbf{(A)_n},$$

where **n** – number of repetitions of the monomer.

Copolymers are macromolecular compounds whose **monomers** have not identical composition. For example:

$$\mathbf{A - C - B - A - C - D - A},$$

where **A**, **B**, **C** and **D** are different monomers. The most important type is the **protein copolymers**, in which the monomers are **aminoacids**.

According to the geometric shape, macromolecular compounds are:

- Linear;
- Branched;
- Spaced.

Depending on the biological role in the human body, macromolecules are:

a) **Biocolloidal** – proteins circulating in plasma and some soluble proteins in the cytoplasm (globulins, fibrinogen);

b) **Structural**, organized in a series of structures, such as muscle protein (actin, myosin), collagen, carotene and so on, which have an extreme importance for the body.

Depending on the informational level macromolecules are classified as high carrier of information (**nucleic acids, enzymes**) and macromolecules with reduced information or lack of information (**polysaccharides**).

Composition, function and electrical properties of proteins

Proteins have the highest portion in cells: 50% and more of the dry weight. The diversity of the living world's protein is 10^{10}–10^{12}. Through the physical and chemical methods it was found that all proteins subjected to examination incorporate C, H, O, N; abundant – S, and other elements such as P, Fe, Zn, Cu.

Usually, the molecular weight is between 5000 and 220000 D, but it can be up to 1000 kD.

By acid hydrolysis, the protein unfolds in simple organic compounds: α-amino acids containing an amino group-NH_2. Amino acids differ by the nature of the radical and side chains (*fig. 2.14*).

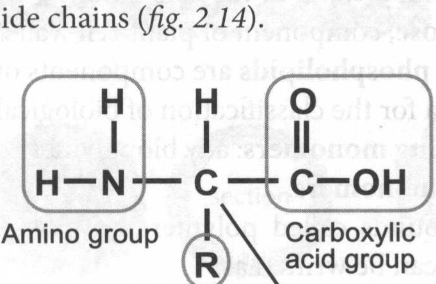


Fig. 2.14.

Proteins are recognized by the structure: **primary**, **secondary**, **tertiary** and **quaternary** (*fig. 2.15*).

Primary structure is a covalently linked polypeptide chain with a specific amino acid sequence.

Secondary structure consists of a single polypeptide chain arranged in a spatial dimension, forming the α-helix structure (folding). Spiral folding is made in accordance with rules established by **Linus Pauling**, for example, coil diameter (10.1 Å) is determined such that the oxygen of the – CO needs to face the hydrogen of the – NH of the next coil. Between O and H forms a hydrogen bond that stabilizes the helix. Step propeller, according to the minimum distance between two points equivalent is 5.12 Å.

Tertiary structure is the three-dimensional spatial structure of a polypeptide chain. For tertiary structure is specific intercuaternare levels, called *heme*, which ensures stability coiled chain.

Quaternary structure involves combinations of several distinct polypeptide chains, forming protein **oligomers**. Chains are called *subunits* or *protomeri components*. Stability and performance characteristics are provided by **links intercuaternara**, **ionic** and **hydrophobic**.

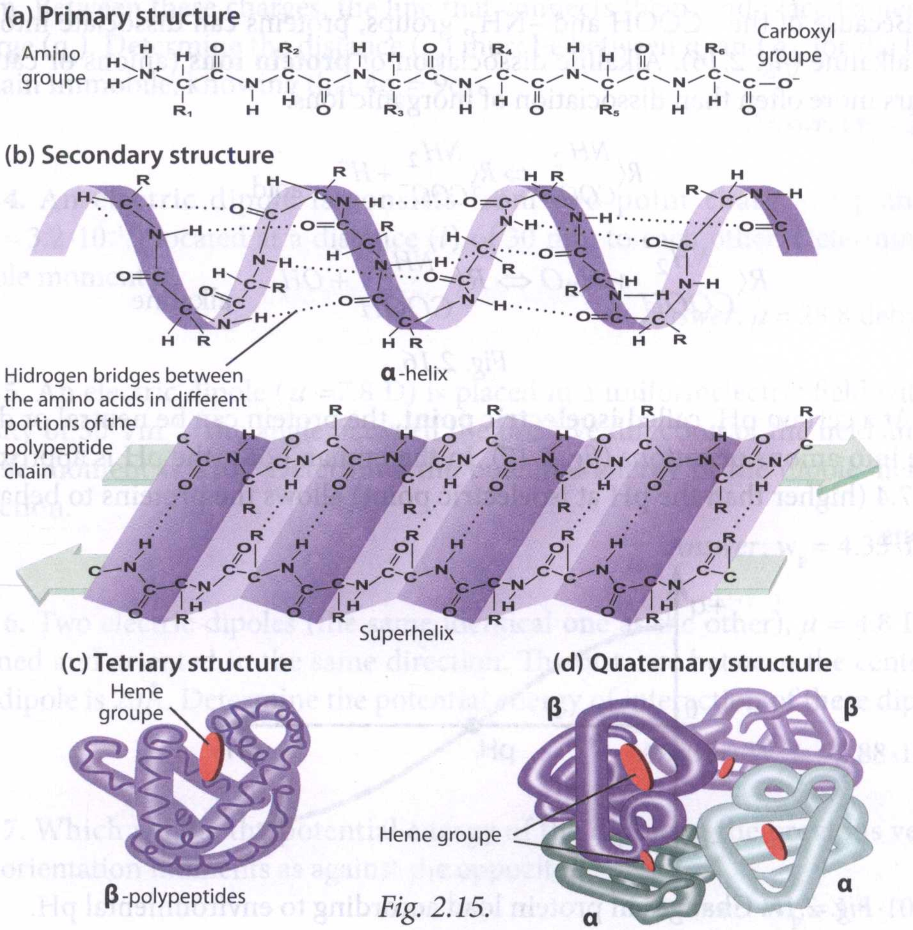


Fig. 2.15.

Spatial structure, tertiary and quaternary play an important role in protein synthesis, antibody production, enzyme reactions, ion channels, pumps, etc. The term conformation refers, usually to the secondary, tertiary and quaternary structure of proteins.

Depending on the conformation, proteins are classified into **fibrillar** and **globular**. Proteins perform very important and diverse functions. They are:

- Structural elements;
- Contractile system components;
- Catalysts (enzymes);
- Transport substances (hormones, toxins, antibodies, ions, etc.);
- Nutrient reserves.

The distortion (dissolution normal structure under the influence of extreme pH or high temperature) of proteins makes them unable to perform their biological functions.

Because of the –COOH and –NH_2, groups, proteins can dissociate into acid and alkaline (*fig. 2.16*). Alkaline dissociation of **protein ions** (anions or cations) occurs more often than dissociation of inorganic ions.

$$R\langle\begin{matrix}NH_2\\COOH\end{matrix} \Leftrightarrow R\langle\begin{matrix}NH_2\\COO^-\end{matrix} + H^+ \qquad \text{acid}$$

$$R\langle\begin{matrix}NH_2\\COOH\end{matrix} + H_2O \Leftrightarrow R\langle\begin{matrix}NH_3^+\\COOH\end{matrix} + OH^- \qquad \text{alkaline}$$

Fig. 2.16.

At a certain pH, called **isoelectric point**, the protein can be neutral or dissociate into anions or cations (*fig. 2.17*). In the human body, the pH is approximately 7.4 (higher than the pH at isoelectric point) allows the proteins to behave as anions.

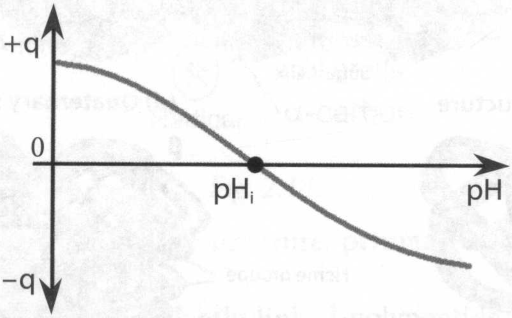


Fig. 2.17. Changes in protein load according to environmental pH.

EXERCIŢII

1. The first electron orbit in the hydrogen atom is circular with a radius of $5.3 \cdot 10^{-2}$ nm. The proton with charge of $+e = 1.6 \cdot 10^{-19}$C is located in the center of the cercle. Determine the potential V created by proton at first electronic orbits.

Answer: V = 27.17 volts.

2. Two point charges $q_1 = 1.6 \cdot 10^{-19}$C and $q_2 = -3.2 \cdot 10^{-19}$C a plased in vacuum at a distans of 0.4 nm each other. Determine the force F, exerced by the charge q_1 on charge q_2.

Answer: F = 2.88 nN.

3. Two positive point charges (q_1 and q_2) a plased in vacuum at distance (r) of 5 nm. Between these charges, the line that connects them, and placed a negative charge (q_3). Determine the distance (r_1) must be between q_1 and q_3, for the last to remain immobile, knowing that $4q_1 = 9q_2$.

Answer: r_1 = 3 nm.

4. An electric dipole is consists from two point charges +q and −q ($|q| = 3.2 \cdot 10^{-19}$) located at a distance (l) of 30 nm, to each other. Determine the dipole moment μ.

Answer: μ = 28.8 deby (D).

5. An electric dipole (μ =7.8 D) is placed in a uniformelectric field with intensity of 50 Vm^{-1}. The angle between the pozitive direction of the field and the dipole moment is 120°. Determine the potential energy of the dipole–field interaction.

Answer: $w_p = 4.33 \cdot 10^{-30}$ j.

6. Two electric dipoles (the same identical one as the other), μ = 4.8 D, are aligned and oriented in the same direction. The distance between the centers of the dipole is 20Å. Determine the potential energy of interaction of these dipoles.

Answer: $W_p = -2.88 \cdot 10^{-22}$ j.

7. Which will be the potential energy of the dipole in the previous year, if the orientation moments as against the oppozite?

Answer: $W_p = 1.8 \cdot 10^{-3}$ eV.

3. WATER. IT'S STRUCTURE AND PROPERTIES

3.1. STRUCTURE AND PROPERTIES OF WATER MOLECULE

The water molecule **H_2O** is formed of one oxygen atom and two hydrogen atoms linked with the oxygen atom by covalent bonds. It was proved by the **X-ray diffraction method** that the water molecule has a triangular shape. It is considered to be an association of two ions of **O^{2-}** and two ions of **H^+** (*fig. 3.1*).

Each link of OH with the length of 0.96 Å represents an electric dipole. Links of OH form between them an angle of 104°, 28′. The resultant dipole moment of the water molecule is 1.84 debye (D).

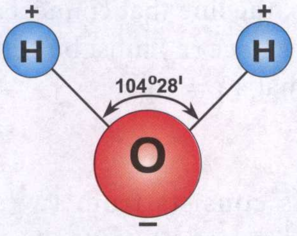


Fig. 3.1. The structure of the water molecule.

Besides the arguments that have been experimentally demonstrated, the fact that the water molecule is a permanent electric dipole, excludes the possibility of a linear arrangement of the component atoms (*fig. 3.2*).

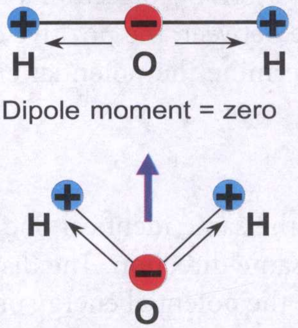


Fig. 3.2.

Another characteristic of the water molecule consists of the fact that it can easily contract with the neighboring molecules bonds (bridges) of hydrogen (*fig. 3.3*).

Hydrogen bonds represent a type of weak intermolecular interactions, which are spread in the macromolecular structures.

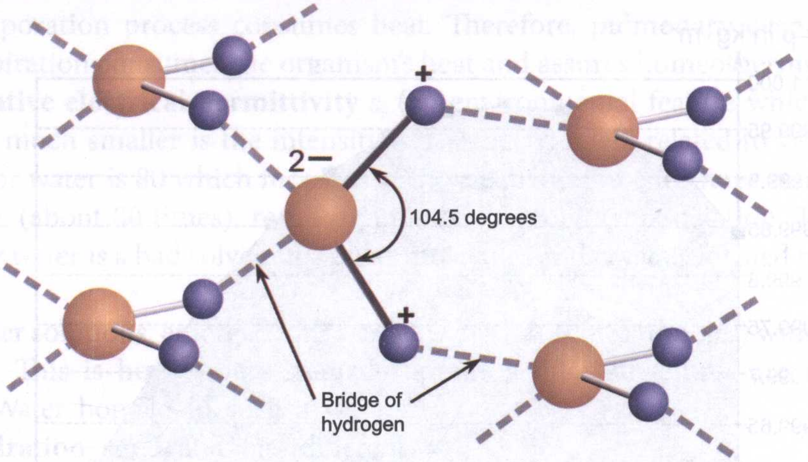


Fig. 3.3.

Regular liquid water is formed of a mixture of independent molecules; monomers H_2O; dimers $(H_2O)_2$; tetramers $(H_2O)_4$ and octamers $(H_2O)_8$; the percentage of each of these fractions depends on the temperature.

In the temperature range of 35–41°C, the dimmers are predominant in water and the chemical activity increases. In such conditions, body cells can easily build linear and spatial structures of water molecules.

This explains the fact that over the time the **warm-blooded** animals have fixated their normal temperature between the limits of the mentioned interval.

3.2. PHYSICAL PROPERTIES OF THE WATER AND THEIR BIOLOGICAL ROLE

The electric dipolar character of the water molecule and its capacity to form hydrogen bonds with the neighboring molecules explains the properties of water in its three states and its behavior in biological environment.

Physical properties of water with remarkable biological importance:

– **Density** (mass per unit volume) **max. et 4°C**, is an exception compared to other substances. Due to this feature, the temperature at the bottom of lakes does not decrease under this rate even during the winter (water with higher density gets down). Ice, which is less dense, raises to the surface so the conditions for further development of submersed life remains favorable.

The diagram of dependence of water's density on the temperature is represented in (*fig. 3.4*).

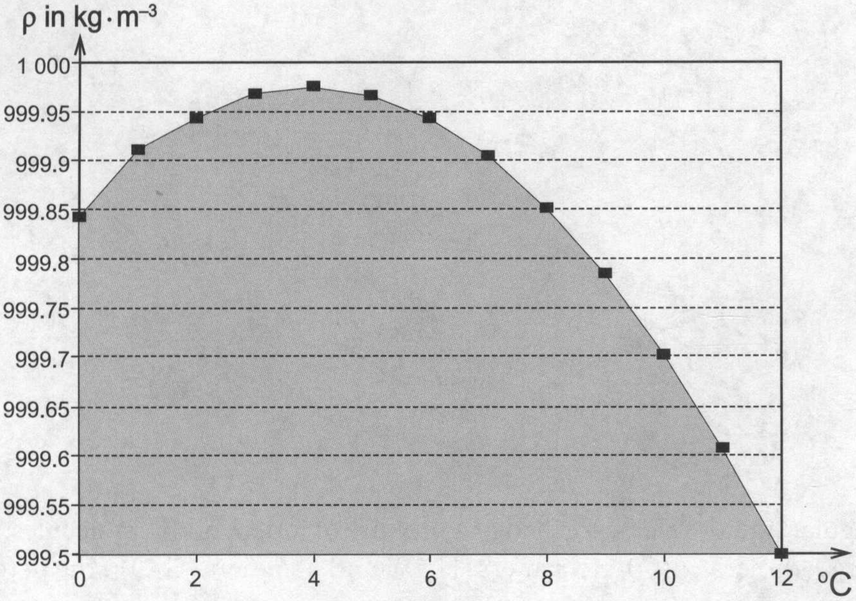


Fig. 3.4.

Between the temperature of 0°C and 4°C, the monomers enter the free spaces of the crystal lattice as interstitial molecules. This phenomenon explains the density growing until it reaches the maximum value. When the temperature is higher than 4°C, these monomers leave through crystal lattice gaps due to Brownian motion, thereby they diminish water's density.

This data draws attention to problems of cryobiology, the freezing and melting of biostructures, which concern researchers due to their non-destructive and functional conservation.

– **Specific heat capacity** (the amount of heat required by a unit in order to change its temperature by 1 degree) is much bigger when compared to other liquid or solid substances ($4.2 \cdot 10^3$ JK^{-1}). This quality can be explained by the fact that the interaction between water dipoles stores a big quantity of intern energy. In body's thermoregulation, the big specific heat contributes to the maintaining of constant temperature during intense muscular stimulation, which can lead to overheating. It's worth mentioning that nowadays biomedical researches still use unsystematic measurement units of thermal energy, which is called **calorie** (1 calorie is equal to 4.18 J).

– **Latent heat specify of vaporization** (the amount of heat required by a mass unit to change its state of matter) is much bigger when compared to other liquids (40.65 $kJmol^{-1}$), due to water's capacity to form hydrogen bonds.

Water evaporation process consumes heat. Therefore, pulmonary evaporation and transpiration consumes the organism's heat and assures homeothermy.

– **Relative electrical permittivity** ε_r (an environmental feature which indicates how much smaller is the intensity of the electric field related to vacuum). Its value for water is 80 which means that the electrical force in water essentially diminishes (about 80 times), resulting in easy electrolytic dissociation. This explains why water is a bad solvent for polar molecules and crystals formed by ionic bonds.

In water solutions, emerged ions attract water molecules (dipoles) with opposite poles. This is how sphere-shaped supramolecular structures are formed (*fig. 3.5*). Water bonded in such a way is called **hydration shell** and has different properties from free water. This process is called **hydration or solvation of ions.**

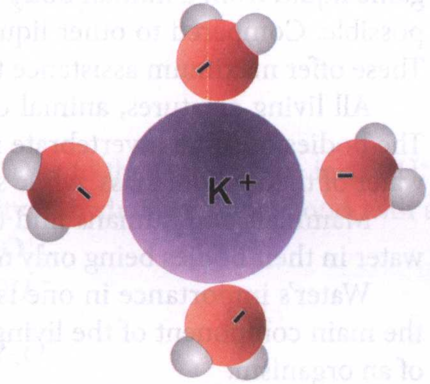


Fig. 3.5.

The number of water molecules which perform hydration, depends on the ion's nature (its charges and sizes).

Surface tension coefficient (σ) (the tension force which returns to the length unity of a contour of the free surface when it interfaces air), of water at 20°C, has a higher value then other liquids ($72.6 \cdot 10^{-3}\,Nm^{-1}$). This fact is important because the surface tension force is involved in a series of important phenomena: air embolism, capillarity, chemotaxis and others.

Here is an example from respiratory system: the tendency of alveoli to collapse at the end of expiration is essentially due to surface tension of water layer which engages alveolar epithelium. The stability of alveoli is insured by **pulmonary surfactant** (a mixture of phospholipids and lipoproteins) from the water-air interface (*fig. 3.6*), which reduces surface tension and maintains the pressure difference in the inside of an alveolus during the respiratory cycle

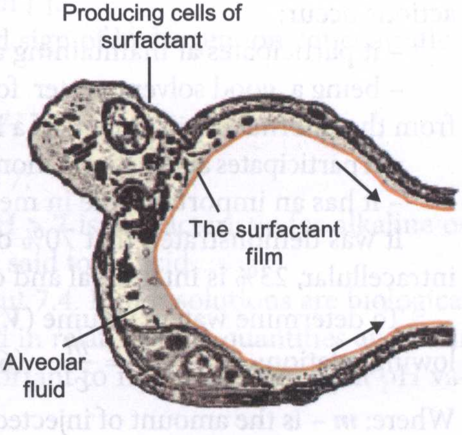


Fig. 3.6.

at an approximate constant value, in this way observing the collapse. The alveoli function correctly when the surfactant layer is present and is synthesized during the whole life by special cells, situated in alveolar walls. The insufficient producing of these, as well as the reduction of tensioactive properties of the surfactant cause different issues of respiratory systems.

3.3. WATER'S BIOLOGICAL ROLE AND WATER BREATHING IN ONE'S ORGANISM

Water – one of the most common substances on our planet – is the only inorganic liquid from a human body without which the vital processes would be impossible. Compared to other liquids, water's physical properties are exceptional. These offer maximum assistance to the occurrence of life phenomena.

All living creatures, animal or vegetal, contain significant amount of water. The bodies of some invertebrate marine creatures (ex. jellyfish) contain 96–97% water of their whole mass, while spores and bacteria contain less than 50%.

Mammals and humans find themselves between those limits, the volume of water in their bodies being only 65–70%.

Water's importance in one is body might be explained by the fact that it is the main component of the living matter and participates in the main processes of an organism.

We mention some of the most important roles of water:

- it is the essential element which regulates osmotic pressure of the three hydrological sectors;
- it represents the environment where many chemical and biochemical reactions occur;
- it participates at maintaining a constant body temperature;
- being a good solvent, water forms a series of solutions which are exchanged from the internal environment of a body into a cell and vice versa;
- it participates at the elimination of the waste products through urinary system;
- it has an important role in mechanical protection of the fetus.

It was demonstrated that 70% of the water contained in a living organism is intracellular, 23% is interstitial and only 7% is circulating water.

To determine water's volume (***V***) in various compartments we can use the following relation: $V = \frac{m}{c}$ (6).

Where: ***m*** – is the amount of injected substance; ***c*** – is the concentration diluted in ***V*** (volume).

To determine the total volume (V_T) of water in an organism it is necessary to inject a substance, which can easily spread in the whole body such as: **antipyrine, hydrogen isotopes** and others.

To find out the plasma volume (the intravascular volume) (V_P) are used the macromolecules which cannot cross the vascular endothelium, such as **radioactive iodinated serum albumin**.

To estimate the extracellular volume (V_{EC}) it is necessary to use a substance that would cross the vascular endothelium but the cell membrane must remain waterproof.

Intracellular (V_{IC}) and interstitial (V_{IS}) volume is determined from the following relations:

$$V_{IC} = V_T - V_{EC} \text{ and } V_{IS} = V_{EC} - V_P \quad (7).$$

3.4. DISSOCIATION OF WATER

Water dissociates into ions of **H⁺** and **OH⁻**. Due to the fact that the ion of H^+, does not have electron shells it has a big capacity of hydration. It can bond with a water molecule and form a **hydronium ion (H_3O^+)**.

$$H_2O \rightarrow H^+ + OH^-$$

$$H_2O + H^+ \rightarrow H_3O^+$$

Due to its mobility, the proton passes from one molecule to another.

The degree of water dissociation is small in pure water at 25°C:

$$\left[H^+\right] = \left[OH^-\right] = 10^{-7} \; mol/l \quad (8).$$

Where the molar concentration is noted with [ ].

The logarithm in base 10 with unaltered sign of hydrogen ion concentration is noted with **pH**:

$$pH = -\lg\left[H^+\right].$$

Therefore, pure water at 25°C has, pH = 7.

When **pH = 7** it's a neutral solution. **pH > 7** is characteristic for alkaline or basic solutions and at **pH < 7** the solution is said to be acid.

The average pH in human's body is about 7.4. Buffer solutions are biological liquids in which beses or acids can be added in relatively big quantities and their pH would not diverge much. They are important to maintain a constant pH values in an organism.

In the following chart is represented the pH of different biological liquids.

The liquid	pH
Blood	7.4
Gastric juice	1.8
Pancreatic juice	8.0
Urine	5.6

The most important buffer solutions in our blood are the serum proteins.

Usually, biochemical reactions are very sensitive to pH's variation and take place in neutral solutions, weak acids and weak basis.

3.5. HEAVY AND TRITIATED WATER

In 1932, **Harold Urey** observed that **the residue of the electrolysis boats** has a bigger density than ordinary water. This helped him to observe **heavy water** which can be found in small amounts in water. Water electrolysis at higher pressure can increase the rate of **heavy water**.

The association of oxygen isotopes (**^{16}O, ^{17}O, ^{18}O**) with hydrogen isotopes (**^{2}H** – deuterium **D** and **^{3}H** – tritium **T**) explains the fact that in natural water exist **18 types** of water molecules, including those which form **heavy water** and **tritiated water**, the las ones being **radioactive**:

H – O – H : H_2O ordinary water
H – O – D
D – O – D : D_2O heavy water
D – O – T
H – O – T
T – O – T : T_2O tritiated water

Physical properties of the heavy water are different from those of ordinary water. The main physical constants of **D_2O** compared to ordinary water can be found in the following chart.

Physical property	H_2O	D_2O
Molar mass ($g \cdot mol^{-1}$)	18	20.0276
Density (grame per cm^3) et 20°C	0.9982	1.1056
Density of maximum temperature	4.0°C	11.6°C
The freezing temperature	0.00°C	3.82°C
The boiling temperature	100.00°C	101.72°C
Surface tension (dynes per cm)	72.75	67.8
Viscosity at 20°C (poise)	$10.09 \cdot 10^{-3}$	$12.06 \cdot 10^{-3}$
Dielectric constant at 0°C	81.5	80.7
The refractive index	1.333	1.328

In biology, we can „label" the water introduced in living organisms – which is one of the methods to study water metabolism. The dosage of heavy water in ordinary water can be determined by measuring the **refractive index**.

Tritiated water (**T_2O**) can be detected based upon the emission of **β** radiation by the tritium according to the following reaction:

$$\mathbf{{}^{3}_{1}H = {}^{2}_{1}H + {}^{1}_{1}H + {}^{0}_{-1}\beta} .$$

Concerning the biological role of heavy water, it was observed that starting with a certain concentration, it slows metabolic processes. Cell division inhibition might be noticed. It causes animal sterility, leads to inhibition of active transport and muscle contractions it generates serious changes in myocardial function which are highlighted in one's electrocardiography and elevates the threshold of excitation.

All these phenomena can be explained if we assume that cell's energetic processes, which are possible in ATP synthesis, the **proton** has an important role that can not be performed by **deuterium**.

EXERCISES

1. What was the initial temperature of 0.8 kilograms of ice, if, its internal energy was raised with $33.6 \cdot 10^3$ J in order to raise its temperature to 0° C? ($c = 2.1 \cdot 10^3\ J \cdot kg^{-1} K^{-1}$).

Answer: $t_i = -20°C$.

2. What volume occupies a certain quantity of water if it contains $6.02 \cdot 10^{28}$ molecules? ($M = 0.018\ kg \cdot mol^{-1}$; $r = 10^3\ kg \cdot m^{-3}$; $N_A = 6.02 \cdot 10^{23}\ mol^{-1}$).

Answer: $V = 1.8\ m^3$.

3. At the boiling point, during 25 minutes, $27 \cdot 10^{-2}$ kg of water evaporated. How many molecules were transformed from liquid state into gaseous state during 1 second? ($M = 0.018\ kg \cdot mol^{-1}$; $N_A = 6.02 \cdot 10^{23}\ mol^{-1}$).

Answer: $N = 6.02 \cdot 10^{22}$ molecules $\cdot\ s^{-1}$.

4. How many times bigger is the energy required to transform into vapors 10 kg of water than the energy required to transform into vapors 20 kg of iron at their boiling point? ($r_a = 22.64 \cdot 10^5\ J \cdot kg^{-1}$; $r_f = 58 \cdot 10^3\ J \cdot kg^{-1}$).

Answer: of 19,5 of times.

5. The temperature of certain amount of water is 80°C. To bring it to the boiling point and then transform it into vapors at a normal pressure were consumed 12 980 kJ of energy. Determine water's mass. ($c = 4.2\cdot10^3\ J\cdot kg^{-1}$; $r = 22.6\cdot10^5\ J\cdot kg^{-1}$).

Answer: $m = 5.5$ kg.

6. The latent heat of vaporization at 37°C is 581 $calories\cdot g^{-1}$. Determine the calorie loss at pulmonary evaporation of 0.6 liters of water.

Answer: $Q = 346.51\cdot10^3$ calories.

7. Determine the extra pressure produced by the surface tension in one drop of fog with the diameter of 3 μm at the temperature of 20°C. ($\sigma = 72.6\cdot10^{-3}\ N\cdot m^{-1}$).

Answer: $\Delta P = 96.8$ kPa.

8. When a drop of pure water is separated from the capillary tube, placed vertically, it weights 40 μg. Determine the diameter of the tube. The experiment takes place at the temperature of 20°C.

Answer: $d \approx 1.76$ mm.

9. In a capillary tube, placed horizontally were introduced 0,3 ml of blood which formed a 12 cm column. What will be the column's height retained in the tube when changing its position to vertical? ($\rho = 1\,060\ kg\cdot m^{-3}$; $\sigma = 58$ mN/m).

Answer: $\Delta h \approx 3.88$ cm.

10. Sometimes the dosage of a medicine is prescribed in drops. How much will change the dosage of the prescribed medication if the temperature varies from $t_1 = 25°C$ to $t_2 = 10°C$? ($\sigma_1 = 71.78$ mN/m and $\sigma_2 = 74.01$ mN/m).

Answer: shall be increased by ≈ 3%.

11. A certain amount of water flowed through a capillary of Oswald viscometer during 60 seconds. The same amount of solution, with double density of the water flowed in 180 s. Viscosity coefficient of the water: $\eta_a = 1$ mPa·s. Determine the viscosity coefficient of the solution.

Answer: $\eta_s = 9\cdot10^{-3}\ Pa\cdot s$.

4. BIOPHYSICS OF DISPERSED SYSTEMS

4.1. DEFINITION AND CLASSIFICATION OF DISPERSED SYSTEMS

We call **dispersed systems** a mixture which is the result of the dispersion of one or several species of molecules called **solute**; among other molecular species called **solvent**. In biology the main solvent is water.

There are several criteria to classify disperse systems.

1. By the grade of dispersion (**$\Delta = 1/d$**) (the opposite of the diameter of dispersed particle). This classification implies a spherical shape of dispersed particles and can't be applied every time.

2. By the number (***n***) of atoms located in the dispersed particle.

3. By the phases. A phase is a homogeneous part of a system, limited by surfaces of the other parts, where there is a sudden variation of physic and chemical properties. The separation area is called **interface**. Depending on the type of the phase, dispersed systems might be:

– **monophasic:** they consist of only one phase, they do not have any discontinuities and are **homogeneous** (similar in all parts of the system);

– **polyphasic** – **heterogeneous**: there are separation spaces between the component parts (fog, aerosols, foam, liquids and gas).

Classification criteria and the names of the dispersed systems are represented in the following chart:

Classification	Molecular solutions	Colloids	Suspensions	Author
By the degree of dispersion	$\Delta > 10^9$	$10^7 < \Delta < 10^9$	$\Delta < 10^7$	Ostwald
After the number of atoms	$n < 10^3$	$10^3 < n < 10^9$	$n > 10^9$	Staudinger
After the phase criterion	monophasic	polyphasic		–

The main classification criterion might be the state of matter of the solvent (the solvit might be liquid, gas or solid). Dispersed systems might be:

– gaseous – the dispersed substance is gas (gas mixtures, air vapors, fog);

– liquid – the dispersed substance is liquid (immiscible liquids, gases);

– solid – the dispersed substance is solid (some alloys).

The particles from dispersed systems might be seen via different methods of optical and electronic microscopy depending on their linear sizes (*fig. 4.1*).

It must be mentioned that biological fluids have a complex behavior and properties combined for all three classes of dispersion. For example, blood is a **solution** for crystalloids (Na, Cl, K), **colloid** (because it contains serum albumins and globulins), **suspension** (due to the figurative elements).

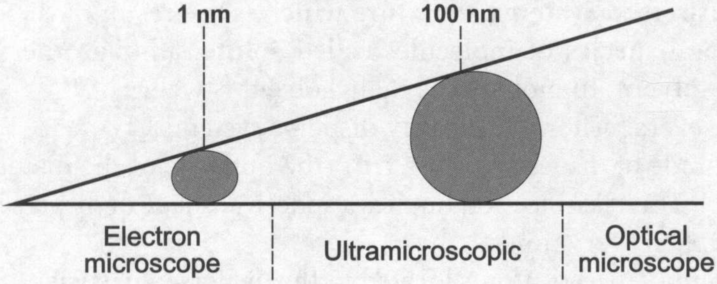


Fig. 4.1. Diameter and visibility of dispersed particles.

4.2. CONCENTRATIONS OF SOLUTIONS

This solution feature can be examined in several ways:

Mass concentration represents solute's mass related to the volume of the solvent. We note it with $\mathbf{c^M}$ and express it with $\mathbf{g \cdot l^{-1}}$ or in SI with $kg \cdot m^{-3}$**.** Even though this method seems to be quite convenient in practice, mass concentration does not allow the comparison between different types of solutions while their properties depend on the number of existent molecules in the solution.

Molar concentration (molarity) represents the number of solvit moles related to the volume unity of solution. It is noted with $\mathbf{C^M}$ and expressed in $\mathbf{mol \cdot l^{-1}}$ or in IS $\mathbf{kmol \cdot m^{-3}}$.

We must notice the fact that one mole is considered to be the amount of substance with a number $N_A = 6.02 \cdot 10^{23}$ of molecules. For example, mass concentration of glucose $M = 180\ g \cdot mol^{-1}$ in ordinary blood $\mathbf{c^M = 1\ g \cdot l^{-1}}$ which corresponds to molar concentration $\mathbf{C^M = 5.5 \cdot 10^{-3}}\ \boldsymbol{mol \cdot l^{-1}}$ which we can determine from the relation below:

$$C^M = \frac{c^M}{M} = \frac{1g \cdot l^{-1}}{180g \cdot mol^{-1}} = 5.5 \cdot 10^{-3}\ mol \cdot l^{-1}.$$

Molar concentration represents the number of moles of the solvit dissolved in a unity of mass of solvent. It is noted with $\mathbf{C^m}$ and expressed in $\mathbf{mol \cdot kg^{-1}}$. Even though it offers multiple advantages in theory, this way to express the concentration in practice is rarely used.

Osmotic concentration represents the number of osmoles in one volume unit of solution, it is noted with $\mathbf{C^O}$ and expressed in $\mathbf{osmoles \cdot l^{-1}}$. One osmole contains N_A mobile particles in form of solution, which can be: molecules, ions, atoms or their aggregates.

In a neutral solution, the number of osmoles is equal with the number of moles. In this way:

$$\mathbf{C^O = C^M osmoles \cdot l^{-1}}.$$

In cases when the solution is electrolytic, the molecules dissociate in ions and the number of osmoles becomes bigger than the number of molecules.

There are frequent cases when the dissociation of molecules in the solvit is partial. **The degree of dissociation** represents the relation between the number of dissociated molecules (**n**) and the total number of solvit molecules ($\mathbf{n_0}$):

$$\alpha = \frac{n}{n_0} \leq 1 .$$

If we know the molar concentration of a electrolytic solution and that $\boldsymbol{\alpha < 1}$ the number of dissociated molecules will be $\boldsymbol{\alpha C^M}$, and the number of the remaining undissociated will be $\mathbf{(1-\alpha)\, C^M}$. So the total number of remaining free particles will be:

$$\nu\alpha C^M + (1-\alpha)C^M .$$

Therefore:

$$C^o = \nu\alpha C^M + (1-\alpha)C^M = [\nu\alpha + (1-\alpha)]C^M \Rightarrow [1+\alpha(\nu-1)]C^M .$$

Where ν is the number of ions formed after the dissociation of one molecule.

The expression $1+\alpha(\nu-1)$ in noted with **i** and called **Van't Hoff ionization factor**. Accordingly:

$$\mathbf{C^O = i\, C^M} \text{osmoles} \cdot l^{-1}.$$

Electrolytes that completely dissociate in water solution, such as $CaCl_2$, are called **strong electrolytes** while those that dissociate only partially, such as CH_3COOH are called **weak electrolytes.**

Obviously, in cases when all solute molecules dissociate ($\alpha = 1$) the numeric ionization coefficient is equal with the number of formed ions at the dissociation of one molecule ($i = \nu$). The relation between osmotic and molecular concentration reappears:

$$\mathbf{C^O = \nu\, C^M} \text{osmoles} \cdot l^{-1}.$$

For example: we have a solution of $\mathbf{CaCl_2}$ with molar concentration 0.2 $mol \cdot l^{-1}$. What would be the osmotic concentration of this solution when $\alpha = 1$ and when $\alpha = 0.9$?

$$C^M = 0.2\ mol/l .$$

The molecule of $\mathbf{CaCl_2}$ dissociates into $\mathbf{Ca^{++}}$ and $\mathbf{2Cl^-}$; $\boldsymbol{\nu = 3}$

a) $\alpha = 1$

$$\boldsymbol{i = 1 + 1(3-1) = 3;} \quad \boldsymbol{C^O = iC^M = 3 \cdot 0.2 = 0.6 \text{ osmoles}/l;}$$

b) $\alpha = \mathbf{0.9}$

$$\boldsymbol{i = 1 + 0.9(3-1) = 2.8;} \quad \boldsymbol{C^O = 2.8 \cdot 0.2 = 0.56 \text{ osmoles}/l.}$$

Ionic concentration (ionization) is denoted with C^I and expressed with $\mathbf{g \cdot ion \cdot} \boldsymbol{l^{-1}}$ or $\mathbf{moles \cdot ion \cdot} \boldsymbol{l^{-1}}$:

$$C^I = \nu\, C^M \text{ – dissociation } 100\%;$$

$$C^I = \alpha \nu\, C^M \text{ – dissociation } \alpha\%.$$

Equivalent concentration (normal) is an equivalent ($\mathbf{E_q}$) wich represents the quantity of a substance, which contains N_A of elementary chargers. The equivalent concentration $\mathbf{C^N}$ represents the number of E_q of solute in one liter of solution.

If one ion has **Z** charge, $\mathbf{C^N = ZC^M}$. For example, the concentration equivalent with the cations in a solution of $\mathbf{10^{-3}\ mol} \cdot \boldsymbol{l^{-1}}$ of $\mathbf{CaCl_2}$ is:

$$\mathbf{C^N = 2 \cdot 10^{-3}\ E_q \cdot \boldsymbol{l}^{-1} = 2\ mE_q \cdot \boldsymbol{l}^{-1}}.$$

Titer – the relation between the mass of the solvit and the mass of the solution. If the units are the same titer can be expressed in percents:

$$\tau = \frac{\text{the solvit mass}}{\text{the solvit mass} + \text{the solvent mass}} \cdot 100\%.$$

4.3. ELECTRICAL PROPERTIES OF IONIC SOLUTIONS

Attraction forces which ensure cohesion between atoms and crystals obey the **Coulomb's law**. In water these forces diminish 80 times, which leads to dissociation of molecules in free ions.

Besides the **dissociation degree** $\boldsymbol{\alpha}$, previously mentioned to appreciate the electrical properties of the solutions there were explained the following measurements:

- **Ionic strength in solution (μ)**

This measurement represents a numeric value which characterizes the degree of electrostatic properties in the solution. It can be defined from this relation:

$$\mu = \frac{1}{2} \sum C_i^I z_i^2 \quad (1).$$

Where: C_i^I – ionisation expressed through **ion moles per liter**;

Z_i – ion charge $\boldsymbol{i}$.

- **Concentration activity (A)**

In real solutions, due to interaction between ions, only one fraction γ of all ions can easily participate at conducting. Its concentration activity is expressed as:

$$A = \gamma C^M \quad (2).$$

Where: γ – activity coefficient; C^M – molar concentration.

When ionic strength is very weak $\mu < \mathbf{0.001}$; $\gamma = \mathbf{1}$.

When $\mu > \mathbf{0.1}$; $\gamma < \mathbf{1}$.

- **Ionic mobility (U)**

On an electric field, with intensity ***E***, a certain force ***F***, influences ion charge ***q*** :

$$\vec{F} = q\,\vec{E} \quad (3).$$

The viscosity of the environment restricts its movement in a solution. The brake force is explained by Stokes law and is also applied for spherical particles:

$$F = -6\pi\eta r v \quad (4).$$

Where: η – viscosity coefficient of the solution; r – ion radius; v – speed.

Making these two forces equal, in module, the movement of the ion becomes uniform, its speed being:

$$\vec{v} = \frac{q}{6\,\pi\eta r}\cdot\vec{E} \quad (5).$$

Through the definition, the term $U = \dfrac{q}{6\,\pi\eta r}$, is considered the mobility of the ion. In consequence, $\vec{v} = U \cdot \vec{E}$ (6).

From relation number (6) it results that mobility is the speed of uniform movement of the ion in a solution, under the influence of a unitary electric field ($E = 1\,Vm^{-1}$).

- **Transport number (t)**

Experimentally determining the mobility of positive ions (**U_+**) and negative ions (**U_-**) for an electrolyte, we can define two relations:

1) $t_+ = \dfrac{I_+}{I} = \dfrac{U_+}{U_+ + U_-}$, is called *the transport number of the cations*;

2) $t_- = \dfrac{I_-}{I} = \dfrac{U_-}{U_+ + U_-}$, is called *the transport number of the anions*.

Where: I – intensity of the whole electric current between electrolytes; I_+ – fraction of the current which corresponds to cations; I_- – fraction of the current which corresponds to anions.

- **Conductibility of a watery ionic solution (χ)**

Electrolytic solutions are good conductors of electric current. The electric resistance of an electrolytic solution with the resistance ρ can be determined from the following relation:

$$R = \rho \frac{l}{S} \ (\Omega); \qquad \rho = \frac{RS}{l} \qquad (7).$$

Where: S – the surface of electrodes in **m²**; l – distance between electrodes in **m**; ρ – resistance in $\Omega \cdot$ **m**.

Electrical conductibility is expressed by the relation:

$$\chi = \frac{1}{\rho} \ \Omega^{-1} \cdot \mathrm{m}^{-1}; \qquad \chi = \frac{l}{RS} \qquad (8).$$

It also can be used the unit **Siemens · m⁻¹**.

Electrical conductibility is proportional to the resistance and can be expressed through the following relation:

$$\chi = FC^{M} Z\alpha (U_{+} + U_{-}) \qquad (9).$$

Where: F – Faraday constant = 96 500 C/mol; C^M – molar concentration of an electrolyte in **mol · m⁻³**.

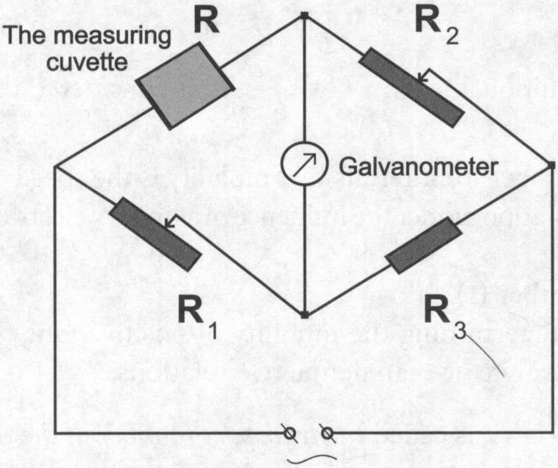


Fig. 4.2.

Conductibility of these solutions can be experimentally determined by using the electric scheme (*fig. 4.2*) called *Kohlrausch bridge*, where: R – resistance of the solution from the measuring cell; R_1 and R_2 – variable resistances; R_3 – the known resistance.

An alternative electric current of 1000 Hz powers the arrangement in order to avoid the polarization of electrodes.

Changing the resistance measurements R_1 and R_2, the bridge equilibrates and the galvanometer (detector) shows the absence of the current.

In these conditions we have the relation: $\frac{R}{R_1} = \frac{R_2}{R_3}$, where from $R = R_1 \frac{R_2}{R_3}$ (10).

Knowing the resistance – R, we can determine the conductivity of the solution:

$$\chi = \frac{l}{RS} \quad (11).$$

Where: l – distance between electrodes from the measuring cell; S – surface of electrodes. The relation l/s is called **the constant of the cuvette.**

EXERCISES

1. A mixture was formed from 10 cm^3 of 10% glucose solution and 60 cm^3 of 25% glucose. Determine the concentration in g/l of the mixture. (The solutions are liquid and considered diluted.)

Answer: $c^M = 228.6$ g/l.

2. There were added 2 liters of water to a glucose solution 7 g/l with the volume of 500 cm^3. Determine the concentration of the new solution, expressed in g/L.

Answer: $c^M = 1.4$ g/l.

3. 35.5 g of Na_2SO_4 were dissolved in 500 cm^3 of water. Determine C^M, C^I, C^N knowing the molar mass of: Na = 23 g/mol; S = 32 g/mol; O = 16 g/mol.

Answer: $C^M = 0.5$ mol/l; $C^I = 1.5$ ion moles/l; $C^N = 1$ mol × E_q/l.

4. A solution of sodium chloride was prepared by dissolving 4.5 g of NaCl in 500 cm^3 of pure water. Determine the ionization and the osmolarity of the solution.

Answer: $C^I = 0.308$ g · ions/l; $C^O = 0.308$ mol · E_q/l.

5. 24 g of weak electrolyte CH_3COOH (M = 60 *g/mol*) were dissolved in one liter of distilled water. Determine the osmolarity of the solution if the coefficient of dissociation (α) is 0.1.

Answer: $C^O = 0.44$ osmoles/l.

6. A solution contains *mEq g/l* ions of Fe^{+++}. Determine mass concentration in *g/l* of iron in this solution. Iron molar mass (M) is 56 *g/mol.*

Answer: $c^M = 37.4 \cdot 10^{-2}$ *g/l.*

7. A solution of $CaCl_2$ has normally 250 *mEq g/l* of Ca^{++} ions. Determine in *g/l* the mass of calcium chloride (**Ca = 40** ***g/mol*****; Cl = 35.5** ***g/mol***).

Answer: $c^M = 13.88$ *g/l.*

8. One liter of glucose solution with concentration $C^M = 2$ *mol/l* was prepared by mixing two solutions with concentrations $C_1^M = 6$ *mol/l* and $C_2^M = 1$ *mol/l.* Calculate the volume of each solution used to obtain the preparation.

Answer: $V_1 = 0.2$ *l*; $V_2 = 0.8$ *l.*

9. Calculate the volume (V) of water which must be added to 1 *l* of glucose solution with the concentration of 0.2 *mol/l*, to obtain a solution with the concentration of 0.03 *mol/l.*

Answer: V ≈ 5.67 *l.*

10. What is the ionic strength in a water solution with 0.025 M of Na_2SO_3?

Answer: $\mu = 0.075$.

11. What is the ionic strength of a solution formed from 1 liter of NaCl with 0.1 M and 0,5 *l* of Na_2SO_4 0.01 M?

Answer: $\mu = 0.077$.

12. What is the ionic strength in a solution, which consists of 0.02 mol of K_2SO_4 and 0.01 mol of Na_2SO_4 in 3 liters of water?

Answer: $\mu = 0.03$.

13. Admitting that an electrolyte AH dissociates into A^- and H^+. Transportation number of H^+ is $t^+ = 0.812$ and the speed of the H^+ ions in a unitary electric field is 32.5 μm/s. Determine the mobility of A^- ions.

Answer: $U_A^- = 7.524$ μm/s.

14. Mass concentration of urea solution is 24 gl^{-1}. Calculate the molar concentration.

Answer: $C^M = 0.40\ mol \cdot l^{-1}$.

15. In a liter of water were dissociated 10 g of potassium iodide (KI) ($M = 176\ g\ mol^{-1}$). Successively determine: a) C^M, considering the diluted solution; b) T (titer) of the solution, expressed in percentage.

Answer: $C^M = 0.06024\ mol \cdot l^{-1}$; $T \approx 1\%$.

16. In 990 g of water were dissolved 30 g of urea ($M = 60\ g/mol^{-1}$). Considering that the relative density of urea compared to water is 1.323 determine osmotic concentration of the solution.

Answer: $C^O = 0.49\ osmol \cdot l^{-1}$.

17. In a jar full with CaCl (completely dissociated) solution were introduced two electrodes with the surface (S) of $1 cm^2$ at a distance (l) of 10 cm one from each other. The resistance between electrodes (R) is 4 kΩ . Knowing the mobility of Ca^{++} and Cl^- ($U_+ = 50$ and $U_- = 70$ in SI), determine the molar concentration of the solution.

Answer: $C^M = 11\ mmol \cdot l^{-1}$.

5. INTRODUCTION TO FLUID BIOMECHANICS

5.1. FLUID STATICS. FLUID PRESSURE

Fluids include liquids and gases. There are **ideal** (perfect) **fluids** and **real fluids**.

There is no internal friction between the molecules of an ideal fluid. Such fluid lacks viscosity and, at a constant volume, all the changes of the fluid occur without energy consumption.

Unlike solids, fluids are composed of mobile molecules and they lack shape.

The molecules of a fluid in gaseous state are compressible, elastic and they occupy the entire space in a container.

The ideal gases obey the Mendeleev-Clapeyron law:

$$PV = \nu RT \quad (1).$$

Where: P – pressure is measured in pascals (Pa); V – volume measured in m^3; T – absolute temperature measured in °K; ν – number of moles of the gas; R – the universal gas constant.

In liquid state the fluid takes the shape of the container it is placed in. Technically, liquids can be considered incompressible.

According to the definition of pressure (*fig. 5.1*), the **static pressure** of a fluid represents the amount of force acting per unit area:

$$P = \frac{F}{S} \quad (2).$$

Where: P – the pressure measured, in pascals (Pa); F – force, in Newtons (N); S - surface are, in m^2.

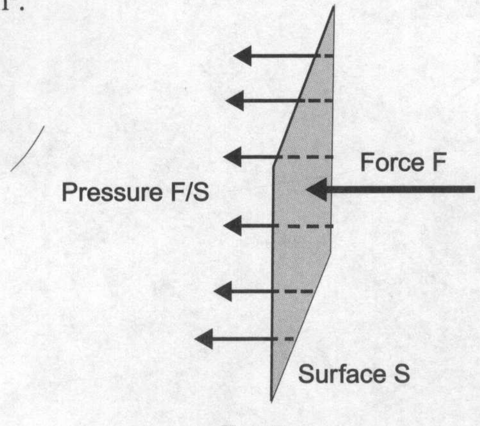


Fig. 5.1.

In fluids, **Pascal's law** states that the pressure, at the same level, is transmitted equally in all the directions throughout the liquid.

There are a lot of applications for the Pascal's principle and one of them is the **hydraulic pressure** whose functioning system uses a piston with a small area **(S_1)** which exerts a small force **(F_1)** on the surface of the liquid (*fig. 5.2*).

According to Pascal's law, the pressure **$p = F_1 / S_1$** is transferred through a connecting tube to a larger cylinder with a bigger surface area **(S_2)**:

$$p = \frac{F_1}{S_1} = \frac{F_2}{S_2} \Rightarrow F_2 = \frac{S_2}{S_1} \cdot F_1 \quad (3).$$

Therefore, the hydraulic press is a machine that amplifies a force with a multiplier equal to the ration between the surface areas of the pistons. The hydraulic pressure is used in dental chairs; hydraulic brakes and the pistons that are pressed correspond to the smaller section.

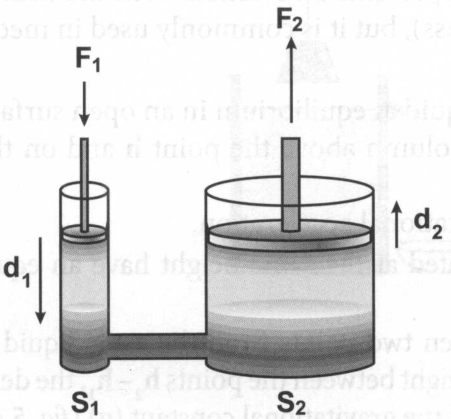


Fig 5.2. Schematic of a hydraulic press.

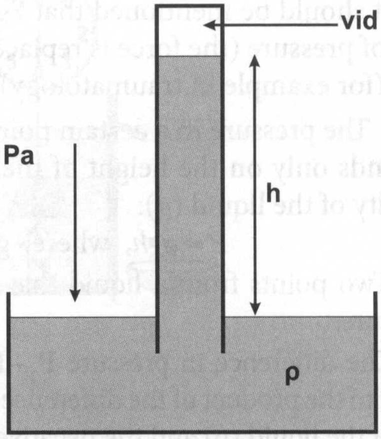


Fig. 5.3.

The **atmospheric pressure** is caused by the weight of the air above the ground. It was discovered by the Italian physicist – **Evangelista Torricelli** (1643). The apparatus used for determining the atmospheric pressure is represented in (*fig. 5.3*).

A glass tube, closed at one end, is filled with mercury, and then it is closed at the other end and after it is tipped into a container with mercury. The mercury in the tube drops to a certain level in comparison to the mercury in the container. This phenomenon occurs when the pressure created in the mercury column balances the atmospheric pressure. The mercury column in the tube is 760 mm at the sea level.

P = 760 mm Hg = 1.013·10^5 Pa is considered to be the normal atmospheric pressure.

The units of measurement for pressure are different for various domains. The most important of them and the conversion coefficients are shown in the table below:

Units of pressure	**Pascal**	**Bar**	**kg · cm^{-2}**	**atmosphere**	**mm Hg**
Pascal	**1**	10^{-5}	$1.02\cdot10^{-5}$	$0.987\cdot10^{-5}$	$0.75\cdot10^{-2}$
Bar	10^5	**1**	1.02	0.987	750
kg · cm^{-2}	$0.980\cdot10^5$	0.980	**1**	0.968	735
atmosphere	$1.013\cdot10^5$	1.013	1.033	**1**	760
mm Hg	133.3	$0.133\cdot10^{-2}$	$1.36\cdot10^{-3}$	$1.315\cdot10^{-3}$	**1**

It should be mentioned that **kg·cm^{-2}** represents a deviation from the definition of pressure (the force is replaced by mass), but it is commonly used in medicine (for example in traumatology).

The pressure in a certain point of a liquid at equilibrium in an open surface depends only on the height of the liquid column above the point **h** and on the density of the liquid (**ρ**):

$P = \rho g h$, where: g – gravitational acceleration.

Two points from a liquid that are situated at the same height have an equal pressure.

The difference in pressure **P$_2$ – P$_1$** between two points from the same liquid is equal to the product of the difference of the height between the points **h$_2$ – h$_1$**, the density of the liquid (ρ) and the negative value of the gravitational constant (g) (*fig. 5.4*):

$$P_2 - P_1 = -\rho g (h_2 - h_1) \quad (4).$$

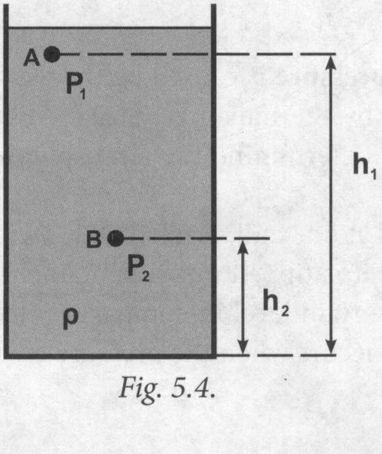


Fig. 5.4.

Archimedes' principle

Archimedes' principle indicates that there is an upward force exerted on a body fully submerged in a fluid that is equal to the weight of the fluid that the body displaces:

$$F = V_c \rho_f g \qquad (5).$$

Archimedes' law also applies to the floating bodies and the weight of the fluid that the body displaces being equal to the weight of the floating body: V_c – the volume of the body (or the volume of the partially submerged part); ρ_f – density of the fluid; g – gravitational acceleration.

The (*fig. 5.5*) represents Archimedes' principle when a body is immerged in water.

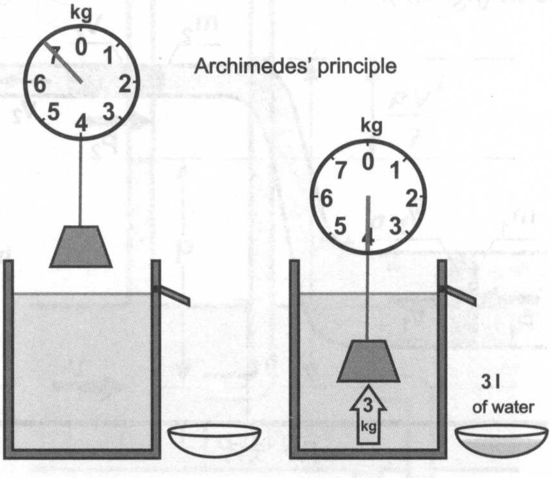


Fig. 5.5.

5.2. FLUID DYNAMICS

Continuity equation

For simplifying, we will consider that fluids are incompressible (the density ρ of the fluid is the same at every point), and the flow – steady (flow velocity at a point does not change over time).

In such conditions, the flow velocity of a fluid in a tube with various areas and horizontally placed (*fig. 5.6*), changes in such way that the product between the velocity (*v*) and the cross-sectional area (*S*) is the same at every point in the tube:

$$S_1 v_1 = S_2 v_2 = \dots = S_n v_n$$

$$S \cdot v = const. \qquad (6).$$

The equation number (6) represents the mathematical expresion for the **continuity equation**.

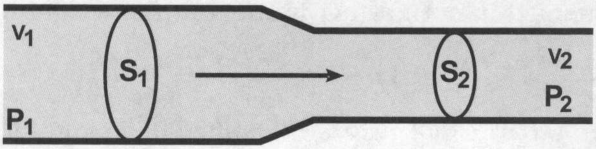


Fig. 5.6.

Bernoulli's equation

We will examine the steady flow of a perfect liquid through a tube with various areas as represented in (*fig. 5.7*).

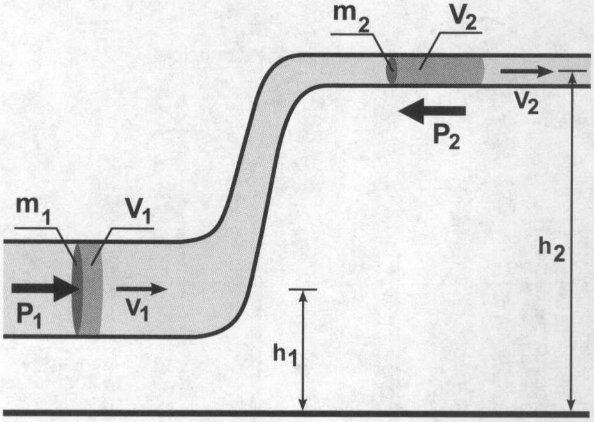


Fig. 5.7.

According to the principle of conservation of energy in mechanics, the total energy of a mass **m** of liquid at any point in the tube must be the same:

$$P_1V+\frac{m\,v_1^2}{2}+mgh_1=P_2V+\frac{m\,v_2^2}{2}+mgh_2 \qquad (7).$$

where: V – volume of the fluid; P – the static pressure at a point in the tube.

Dividing the expression (7) by the volume we obtain:

$$P_1+\frac{\rho\,v_1^2}{2}+\rho gh_1=P_2+\frac{\rho\,v_2^2}{2}+\rho gh_2 \qquad (8).$$

From expression number (8) results that at every point in the tube we get the expression:

$$P+\frac{\rho\,v^2}{2}+\rho gh=const. \qquad (9).$$

Where: P – the static pressure at that point; ρ – the density of the fluid; v – flow velocity; h – the height of the point in the fluid.

This expression represents the total pressure of the liquid and is called **Bernoulli's equation.**

$\frac{1}{2}\rho v^2$ – **dynamic pressure;** ρgh – **hydrostatic pressure.**

When the tube where the liquid flows is horizontally placed (***h*** = ***const.***), we obtain:

$$P+\frac{1}{2}\rho v^{2}=const. \qquad (10).$$

The static, dynamic and total pressure can be easily measured with a simple apparatus called **Pitto's double lumen catheter**.

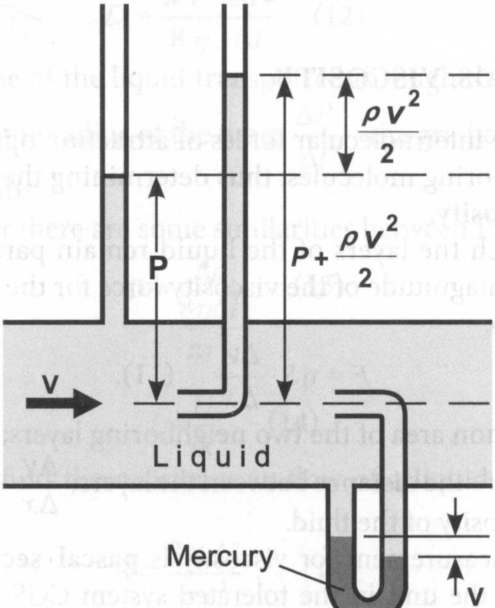


Fig. 5.8.

The manometer tube, parallel to the direction of flow section measures the static pressure and when perpendicular to the direction of flow, it measures the total pressure. As we noticed from (*fig. 5.8*), the difference between the two pressures represents the dynamic pressure.

This apparatus is often used to determine the flow velocity because it has a mercury manometer with the scale of the velocity units on it.

Bernoulli's law implies that in the narrow portions of the circular shaft, where the flow velocity has a big value, the lateral pressure is minimal (*fig. 5.9*). Thus, in cases of artherosclerosis regression, the phenomenon leads to emphasising the closure of the vascular lumen in the region of obstruction.

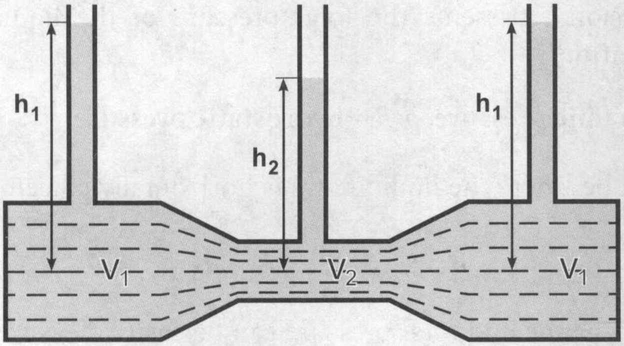


Fig. 5.9.

5.3. REAL FLUIDS. VISCOSITY

In real fluids, the intermolecular forces of attraction oppose the relative movement of the neighboring molecules, thus determining the appearance of a friction force called **viscosity.**

The flow in which the layers of the liquid remain parallel to each other is called **laminar**. The magnitude of the viscosity force for the laminar flow is given by Newton's equation:

$$F = \eta S \cdot \frac{\Delta v}{\Delta x} \qquad (11).$$

Where: S – the common area of the two neighboring layers; Δv – the speed difference of the layers; Δx – the distance between the layers; $\frac{\Delta v}{\Delta x}$ – gradient of velocity; η – the dynamic viscosity of the fluid.

The **SI** unit of measurement for viscosity is **pascal·secund (Pa·s)**, which is 10 times bigger than the unit in the tolerated system CGS called **poise (P)**. The submultiple of poise is called **centipoise** (1 cP = 0.01 P) and it is commonly used in the medical practice.

The dynamic viscosity depends on the fluid and on the temperature (it decreases with the increase of temperature). You can find the values of the dynamic viscosity of some biological liquids in the table below.

Liquid	**Temperature (°C)**	**Viscosity (poise)**
Water	**20**	**0.01**
Glycerin	**20**	**8.3**
Mercury	**20**	**0.0155**
Air	**20**	**0.00018**
Blood	**37**	**0.04**
Urine	**20**	**0.0102–0.0114**

The liquids to which equation number (12) can be applied are called **Newtonian**. Many liquids of medical importance (water, blood serum, cerebrospinal fluid, urine) are **Newtonian liquids**. There still are other liquids, for example macromolecular solutions like blood, which do not act according to equation number 12. Therefore, these liquids are called **non-Newtonian**.

Poiseuille's law

Poiseuille's law applies for the laminar flow of a real liquid through a horizontally placed cylindrical tube, and is mathematically expressed by the following expression:

$$D=\frac{\pi r^4}{8\eta}\frac{\Delta P}{\Delta l} \qquad (12).$$

Where: D – the volume of the liquid transported through the tube in an interval of time(the debit); r – the radius of the tube; $\frac{\Delta P}{\Delta l}$ – the gradient of pressure of the Δl is segment of the tube.

We can notice that there are some similarities between Poiseuille's law:

$$D=\frac{\Delta p}{\frac{8\eta\Delta l}{\pi r^4}} \qquad (13).$$

and Ohm's law:

$$I=\frac{U}{R} \qquad (14).$$

This way we can find the resistance to flow of a liquid through cylindrical tubes:

$$R=\frac{8\eta\,\Delta l}{\pi r^4} \qquad (15).$$

In cases when the tubes are connected in series, the resistance to flow has the following expression:

$$R_{total}=R_1+R_2+....+R_n \qquad (16).$$

In cases when the tubes are connected in parallel, we get this another expression:

$$\frac{1}{R_{total}}=\frac{1}{R_1}+\frac{1}{R_2}+...+\frac{1}{R_n} \qquad (17).$$

Stokes' law

A fluid shows its viscosity not only when it flows through tubes, but also when a body flows through a fluid. **Stokes' law** applies only to spherical shaped bodies with a uniform motion through a fluid (v = const.). The drag force (F) opposing the movement of the body through the liquid is directly proportional to the dynamic viscosity of the fluid (η) al fluidului, radius of the body (r) and velocity of the body (v):

$$F = 6\pi \eta r v \quad (18).$$

We know that a spherical body in a free fall, in a viscous fluid reaches a speed limit ($v_{\lim}$), for which the sum of the drag force produced by viscosity and Archimedes' force is equal to the weight of the body. If ρ is the density of the body, and ρ' – the density of the fluid, we obtain:

$$4\pi r^3 \rho' g/3 + 6\pi\eta r v_{\lim} = 4\pi r^3 \rho g/3$$

$$v_{\lim} = 2r^2 g(\rho - \rho')/9\eta \qquad (19).$$

Measuring the speed limit we can determine the value for the dynamic viscosity of the fluid.

A similar expression, but with a different numerical coefficient, can be obtained for non-spherical bodies. For example, in biology speed limit is replaced by **sedimentation rate**. The experiments, which explain the sedimentation phenomenon, can give useful information about extremely small particles (for example, erythrocytes).

Haematocrit

The volume percentage of the red blood cells in blood is called haematocrit. A healthy person has a haematocrit value between 45 and 50%. The dependency of dynamic viscosity of haematocrit (H) is shown in (*fig. 5.10*).

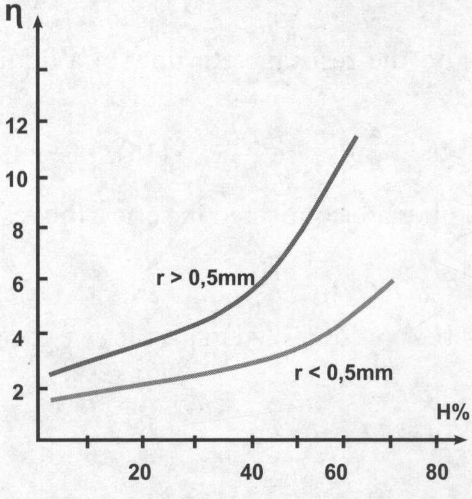


Fig. 5.10.

From figure 5.10 we can notice and interesting phenomenon: the value of dynamic viscosity depends on the radius of the used capillary. This phenomenon is known as the **Fahraeus–Lindqvist effect**:

$$\eta_r = \frac{\eta_\infty}{\left(1+\dfrac{d}{r}\right)^2} \quad (20).$$

Where: η_∞ – dynamic viscosity measured with a viscometer with a large capillary radius; d – diameter of the erythrocytes; r – radius of the capillary.

The dynamic viscosity of the blood flowing through capillaries (r – 2.5 μm) **has a minimal value and mechanical work performed by the heart to pump blood through the capillaries of the vascular system.**

Two flow type. Reynolds number

The flow of a viscous fluid flowing through a tube depends on viscosity and density of the fluid and on the flow velocity and diameter of the tube.

Laminar flow (lamina – a thin sheet): different layers of molecules flow in the same direction (parallel to each other) without mixing. Flow velocity decreases from inside (maximum speed) to the borders of the wall (where the speed is equal to 0).

Turbulent flow: the molecules move on different directions, the substance mixes; *eddies* form inside the fluid. They produce an increase to resistance to flow. The particles may have a high speed near the walls of the container too.

Reynolds's number : $R_e = \rho \text{v} \dfrac{d}{\eta}$ (21).

Where: ρ – the density of the fluid; v – velocity; d – diameter of the container; η – viscosity of the fluid.

The flow is **laminar** if **Re < 2000** and becomes **turbulent** if **Re > 3000**. For the interval between **2000 < Re < 3000** there is a **transitional flow regime**, the flow is **unstable** and it can easily change from a regime to another (*fig. 5.11*).

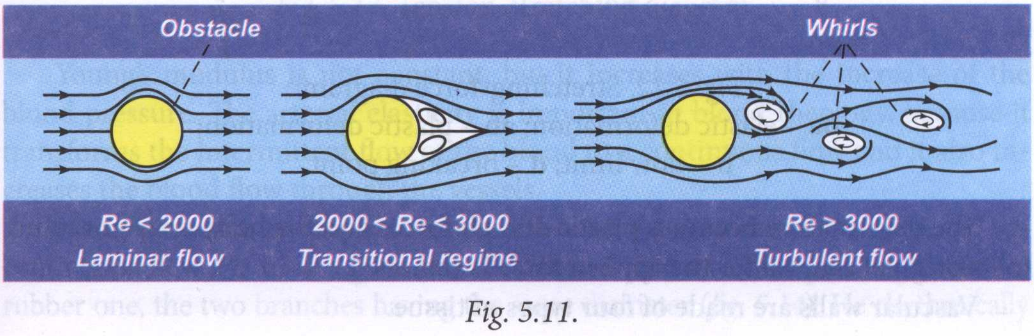


Fig. 5.11.

5.4. VASCULAR WALL ELASTICITY

Structure and elasticity of the wall of blood vessels

Elasticity is the property of a body to change its shape when a force is applied and then return to the initial shape after the force is removed.

Hooke's law applies to the homogeneous elastic bodies, so as the relative elongation $\frac{\Delta l}{l}$ is proportional to the tension $\frac{F}{S}$:

$$\frac{\Delta l}{l} = \frac{1}{E} \cdot \frac{F}{S} \qquad (22).$$

Where: Δl – elongation; l – the initial length; E – Young's modulus, a constant that characterizes the body; F – the force that produces elongation; S – cross-sectional area of the body.

Young's elastic modulus represents the tension that 100% elongation ($\Delta l = l$) would produce.

Hook's law applies only for elastic deformations (*fig. 5.12*).

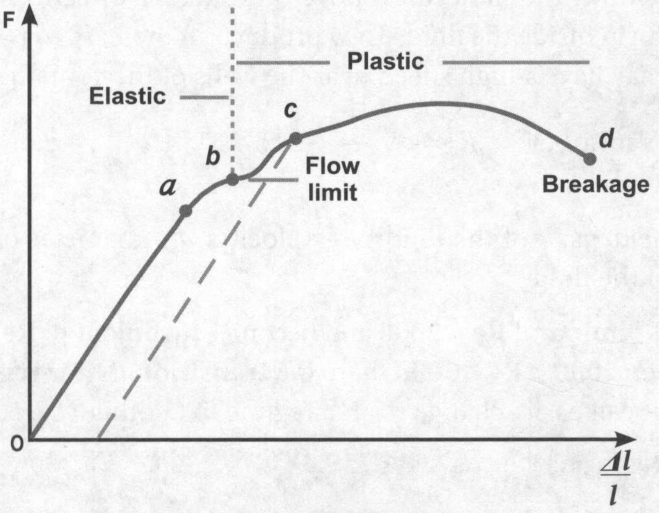


Fig. 5.12. Stretching-force diagram:
oa – elastic deformation; **ab** – plastic deformation;
b – flow limit; **d** – breaking point.

The deformation becomes plastic at high forces and the body does not returt to its original shape after the applied force is removed.

Vascular walls are made of four types of tissue.

1. Endothelium is a layer of flattened cells that lines the interior surface of the wall creating a smooth surface. It has a selective permeability over different substances.

2. The elastic fibers have an elasticity modulus of $3 \cdot 10^5\ N \cdot m^{-2}$. They can be easily stretched and they create an elastic tension on the vascular wall that confers a minimum resistance on the distention produced by the blood pressure.

3. Collagenous fibers have an elasticity modulus of $10^8\ N \cdot m^{-2}$. They are more resistant to stretching that the elastic fibers and give the blood vessel resistance to high pressures. They have a folded structure so they are involved only after a certain stretch (when loops unfold).

4. Smooth muscle fibers prevail as the arterial diameter becomes smaller and is the most developed in arterioles. The smooth muscle tissue creates an active tension that is physiologically controlled. Their constriction changes the radius of the vessel and the debit of the blood that flows through it.

Hooke's law cannot be applied to the blood vessels because they do not represent a homogenous body. The graph line of the dependence of the stretching force on the elongation is represented by the *tension-extension curve* which is not linear (*fig. 5.13*).

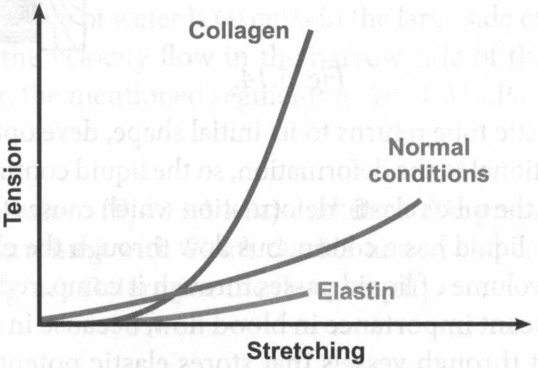


Fig. 5.13. Tension-stretching diagram.

Young's modulus is not constant, but it increases with the increase of the blood pressure. The arterial elasticity is important in blood rheology, because it transforms the intermittent flow of the blood to a continuous flow and it also increases the blood flow through the vessels.

Marey's experiment emphasized the difference between the intermittent and the continuous flow. He took a glass tube detached it into two branches, a glass and a rubber one, the two branches having the same diameter (*fig. 5.14*). He rhythmically pumped a water current through an extremity of the tube. He noticed an intermittent flow through the glass branch of the tube, because the atmospheric pressure was apposing to flow with every pause in flow. The flow was continuous in the rubber branch of the tube, but it has a low velocity. After measuring the volume of the liquid that passed through both of the branches at an equal time interval, he noticed that more liquid flew through the elastic branch even if the diameters of the branches were equal.

The elasticity of the rubber tube explains this phenomenon. The liquid pressure that comes from a tube with a constant flow acts on the liquid from the tube, pushing it forward, and of the elastic walls of the tube, causing an elastic deformation.

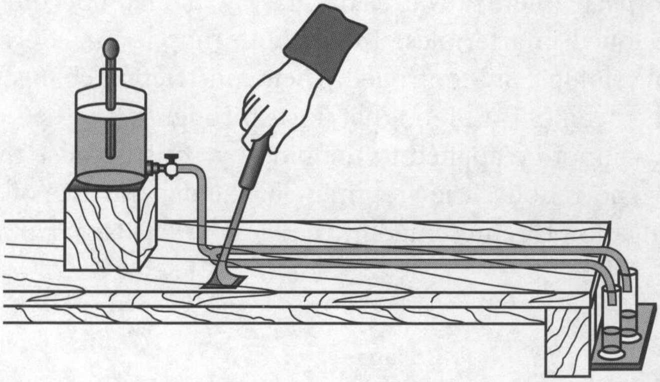


Fig. 5.14.

The deformed elastic tube returns to its initial shape, developing an elastic force that is directly proportional to the deformation, so the liquid continues to flow through the space added by the tube's elastic deformation which causes the tube to increase its diameter. Thus, the liquid has a continuous flow through the elastic tube, a lower velocity and a higher volume of liquid passes through it compared to the glass tube.

This is of a significant importance in blood flow, because in the pulsed regime imposed by the heart through vessels that stores elastic potential energy during diastole, providing a greater flow of blood than if the vessels would have rigid walls.

EXERCISES

1. Determine the pressure exerted on the floor by a man that weights 80 kg and has a shoe sole area of 400 cm^2.

Answer: 19.6 kPa.

2. Determine the pressure exerted by an aluminum cube with a side of 0.2 m ap of one horizontal souport (density of aluminum – $\rho = 2.7 \cdot 10^3$ kg/m^3).

Answer: $5.4 \cdot 10^3$ Pa.

3. How many times will Archimedes' force increase if a body is transferred from oil with density $\rho_u = 900$ kg/m^3 into sulfuric acid with density $\rho_a = 1800$ kg/m^3?

Answer: of 2 times.

4. The pressure in the tire of a truck at a temperature 7°C is $4 \cdot 10^5$ Pa. How much the pressure will increase if the temperature inside the tire becomes 27°C after the truck begins to move?

Answer: with $0.286 \cdot 10^5$ Pa.

5. The flow velocity of water through one of the horizontally placed tube sections is 5 cm/s. Determine the flow velocity of water through that side of the tube that has a diameter 3 times smaller.

Answer: 0.54 ms^{-1}.

6. The flow velocity of water is 60 cm/s in the large side of a horizontally placed tube. Determine the velocity flow in the narrow side of the tube if the pressure difference between the mentioned segments is $\Delta P = 1.33$ kPa.

Answer: ~1.73 m·s^{-1}.

7. The total pressure ($P_t = 13.3$ kPa) and the static pressure ($P_s = 2.66$ kPa) of a liquid were determined with Pitto's double lumen catheter. The density of the liquid (ρ) is 820 kg/m^3. Determine the flow velocity.

Answer: ~5.1 ms^{-1}.

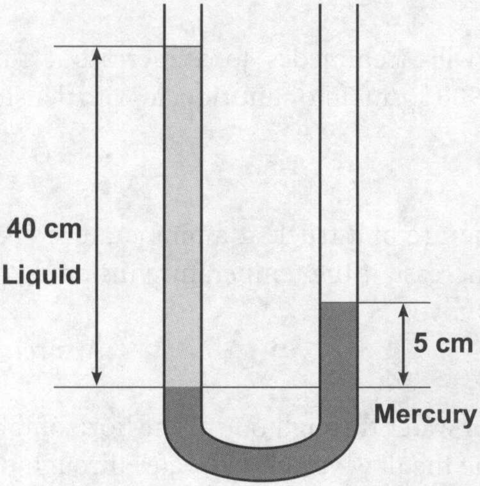
8. The liquid from a large container flows down through a narrow region. Determine the flow velocity when the height (h) of the liquid in the container is 1 m.

Answer: ~4.43 ms^{-1}.

9. It took 60 s for water to flow through the capillary of a viscometer, and 180 s for a 3 times higher density solution. The dynamic viscosity of water (η_0) is 1 Pa·s. Determine the viscosity of the solution.

Answer: $9 \cdot 10^{-3}$ Pa·s.

10. A U-tube is filled with mercury ($\rho_m = 13.6\ \text{g} \cdot \text{cm}^{-3}$) and another with liquid. The free surface of the liquids is set 5 and 40 cm above the separation surface. Determine the density of the unknown liquid.



Answer: $1.7\ \text{g} \cdot \text{cm}^{-3}$.

11. A body made of gold and copper alloy has a mass of 1 kg. When introduced into water its weight is 9.22 N. There are given: $\rho_{Cu} = 9 \cdot 10^3\ \text{kg/m}^3$; $\rho_{Au} = 2 \cdot 10^4\ \text{kg} \cdot \text{m}^{-3}$; $\rho_{water} = 10^3\ \text{kg/m}^3$. Determine the volume of gold.

Answer: $V_{Au} = 4.2 \cdot 10^{-5}\ \text{m}^3$.

12. Determine the ration between the volume of an iceberg above the water level and its total volume (V_t) knowing the density of water ($\rho_a = 1000\ \text{kg} \cdot \text{m}^{-3}$) and the density of ice ($\rho_g = 900\ \text{kg} \cdot \text{m}^{-3}$).

Answer: 10^{-1}.

6. ELEMENTS OF HEMODYNAMICS

Hemodynamics studies physical phenomena of blood circulation (heart mechanics and hidrodynamics of blood flowing through elastic vessels), the gadgets, models and experimental devices used for its purpose. The studies of blood flow uses mechanical models due to series of analogies which exist between the function of the heart and the function of a pump, between arteries and elastic vessels.

6.1. HEART'S ROLE AS A PUMP

The heart is a hollow, muscular organ that pumps blood (a Newtonian fluid) in the whole body by rhythmic contractions (due to cardiac cycle) in blood vessels that have various diameters and whose walls are not rigid and partially elastic. The heart makes about 60–100 beats per minute and 100 000 beats per day. Heart beats might be accelerated by muscular activity and by the higher temperature of the body.

In the heart, the automatist property appears with action potential and is assured by the specific structure of the heart and especially by the **pacemaker cells**. In these cells the excitation appears spontaneously and form the **sinoatrial node**, which is also called the **ruler of heart rhythm**.

These structures are small and have a strict position. In warm-blooded animals' hearts the pacemaker cells are situated in the wall of right atrium, in frog's hearts they are situated in the left sinus.

From this node the excitation is transmitted further to the working myocardium of atriums and later ventricles through nodes and fascicules that together form the **excito-conductive system** of the heart. The **pacemaker cells** are connected by „gap" junctions, which guarantee maximal speed of propagation of excitation. Typical cells of working heart also conduct nerve impulses with a lower speed. In physiological conditions when the conduction system of the heart is a whole, excitation begins with the conductor of cardiac rhythm, goes through the excite-conductive system and reaches final fibers, whence the working myocardium of the heart.

The most important role of the heart consists in blood expulsion in circulation by passive opening and closing of the valves. The heart consists of two pumps (*fig. 6.1*), connected by pulmonary and systemic circulation.

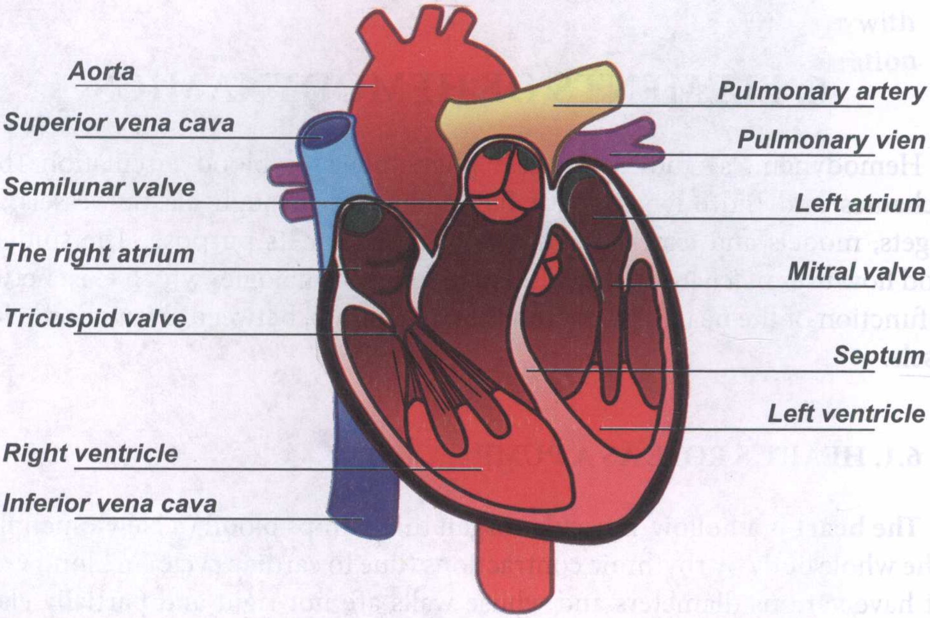


Fig. 6.1.

Right ventricle pump has the function to pump the deoxygenated blood gathered from the whole body to lungs (pulmonary circulation).

Left ventricle pump gathers the oxygenated blood from lungs and pumps it in our body (systemic circulation).

Each part of our heart has two sets of valves, which normally impose the blood flow in one direction. The two heart pumps have each two chambers: **atrium** is a collector of blood brought by veins and **ventricle** pumps blood in arteries. **Septum** is the wall of tissue that separates the atria from the ventricles, preventing the blood to get in other **ventricles/atria**. Heart muscles determine the tightness of these pumps.

Valves motion is set by the pressure difference between atria, ventricles and blood vessels they prevent the blood to flow in the wrong direction. Cardiac muscles assure variation of heart volume and blood pressure, as well as the necessary functional energy required by biophysical and chemo-mechanical processes in myocardium.

Phases of the cardiac cycle

Heart's activity as a pump can be appreciated with the help of cardiac output, which represents the blood volume ejected by each ventricle in one minute. It is equal to the volume of pumped blood by one ventricle at each heartbeat (stroke volume), multiplied by heart rate. **Stroke volume** of each ventricle is approximately 70 ml and the normal heart rate is 70–75 beats per minute; this way cardiac output at rest is 5 l/min. Heart must set in motion 4 l of blood in rest and up to 20 l during exercise. While sleeping cardiac output decreases but it increases during febrile states, pregnancy and at high altitude.

Each heartbeat consists of a sequence of events that represent the cardiac cycle. This consists of three phases (*fig. 6.2*): atrial systole, ventricular systole and diastole.

Atrial systole – represents the contraction of both atria followed by blood inflow in ventricles. When atria are completely empty, atrioventricular valves close and prevent the blood to return in atria.

Ventricular systole – represents the contraction of ventricles and ejection from blood ventricles, which enter the circulatory system. When ventricles are completely empty, pulmonary and aortaic valve close.

Diastole – represents the relaxation of atria and ventricles followed by atria refill.

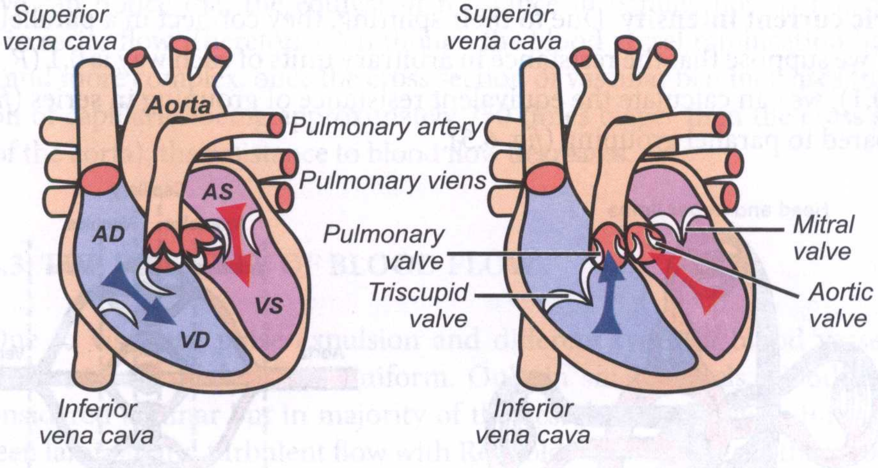


Fig. 6.2.

The specific sounds produced during the heartbeat are a consequence of the closure of the atrioventricular valves and atria and are audible with a stethoscope. According to conservation of energy law, heart's mechanical work can be found as other forms of energy:

- in potential energy of blood (effective pressure on vessel walls corresponds to it);
- in kinetic energy of blood which measures blood motion;
- in blood warming as a consequence to friction between blood layers.

Mechanical work generated by heart in systole is partially accumulated as elastic potential energy of arterial walls and is later transferred to blood during diastole. Because the arteries have elastic walls in conditions of pulsating regime in which heart works, they allow the blood flow during diastole; this way the output is much bigger than it would be in vessels with inflexible walls (see Marey experiment).

6.2. GENERAL SCHEME OF VASCULAR BED

In figure 6.3, is represented a general scheme of vascular bed. From **aorta**, where pressure is the highest (average pressure being 100 mmHg) the blood flows towards the place with lowest pressure – **vena cava** (average pressure being 10 mm Hg).

It is possible to make an analogy with electric current which moves from high potential energy to low potential energy, **blood flow** represents the **equivalent of electric current intensity**. Due to their splitting, they connect in a parallel way.

If we suppose that the resistance in arbitrary units of each way is 0.1 ($R_1 = R_2 = R_3 = 0.1$), we can calculate the equivalent resistance of grouping in series (*fig. 6.4*) compared to parallel grouping (*fig. 6.5*).

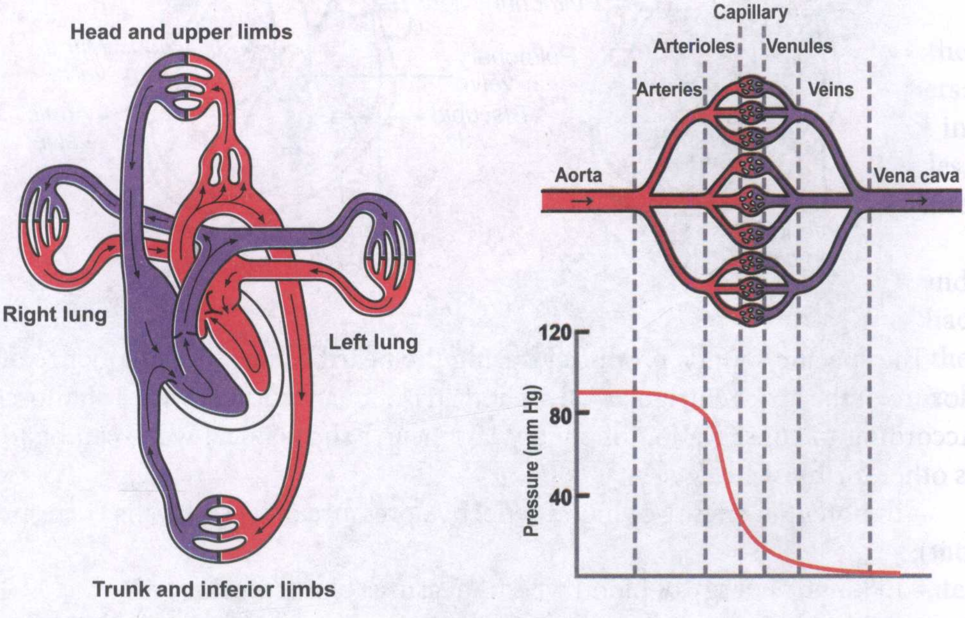


Fig. 6.3.

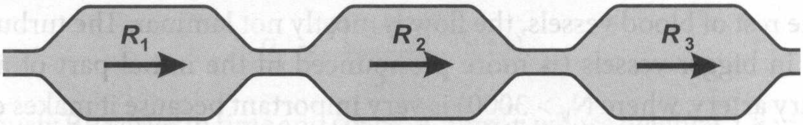


Fig. 6.4. Grouping of capillaries in series.

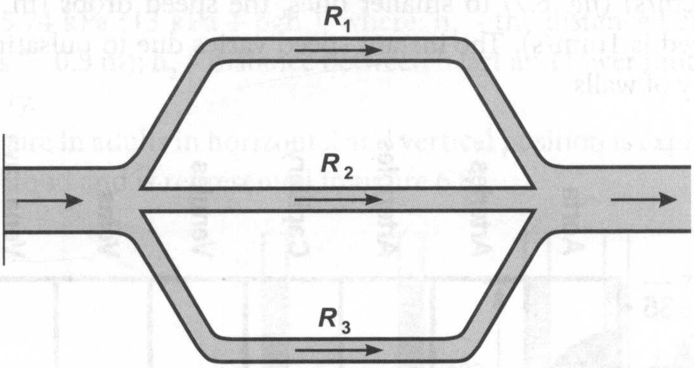


Fig. 6.5. Grouping of capillaries in parallel.

We can notice that the equivalent resistance in parallel flow is much lower than in series flow. Therefore even though the blood vessel ramification becomes more and more complex, once the cross section of vascular bed increases (the total section of capillaries being approximately 750 times bigger than the cross section area of the aorta), the resistance to blood flow decreases.

6.3. THE VELOCITY OF BLOOD FLOW

Due to viscosity, pulse expulsion and different types of blood vessels, the blood flow in our body is not uniform. Only in small vessels, blood flow can be considered laminar but in majority of the vessels, blood flow is intermediate between laminar and turbulent flow with Reynolds number bigger than 2 000 and smaller than 3 000.

Through capillaries, which have smaller diameter than erythrocytes, there is elastic deformation of these so they move one by one with very low speed, being driven by plasma (*fig.* 6.6).

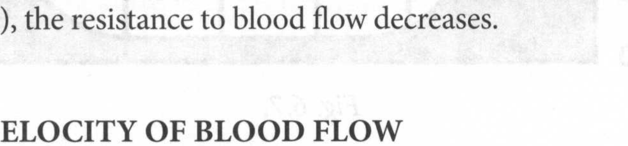


Fig. 6.6.

In the rest of blood vessels, the flow is mostly not laminar. The turbulent flow of blood in bigger vessels (is more pronounced in the initial part of aorta and pulmonary artery, where $N_R > 3000$) is very important because it makes easier the exchange between the fluid and vessel walls. Moving from big vessels (medium speed – 40 cm/s) (*fig. 6.7*) to smaller ones, the speed drops (in capillaries the medium speed is 1mm/s). The instant speed varies due to pulsating regime and deformability of walls.

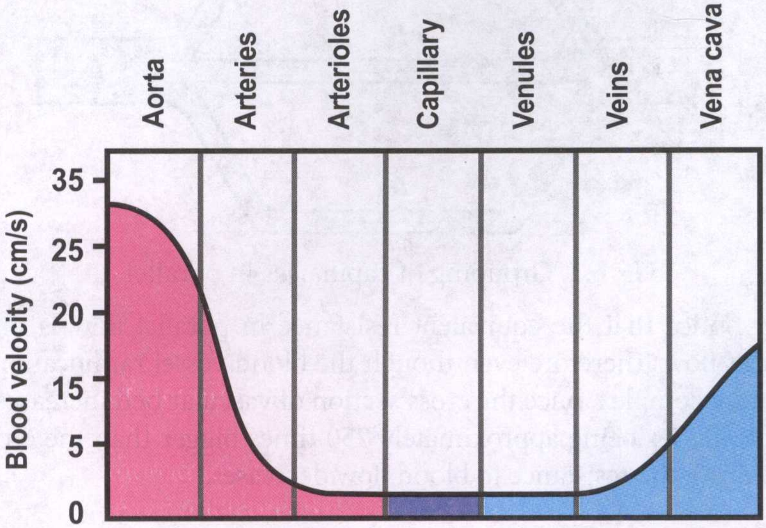


Fig. 6.7.

Applying the continuity equation in blood flow, we can write:

$$S_{total\text{-}capillary} \cdot V_{capillary} = S_{aorta} \cdot V_{aorta}\ .$$

6.4. BLOOD PRESSURE

Blood pressure (**BP**) is the surplus of pressure in the arterial blood vessel reported to atmospheric pressure and measured at the heart level.

Caused by the pulsatile character of the vascular system, BP varies in time between the maximal value of **systole** (**P_s**) and the minimal of **diastole** (**P_d**). A healthy grown-up's systole pressure is approximately 120 mm Hg (15 kPa), and the diastolic is approximately 80 mm Hg (10 kPa). The average value of BP can be determined from the following relation:

$$P_m = \frac{P_s + 2P_d}{3} \qquad (3).$$

Obviously BP varies in time and space. A person whose height is 1,8 m, average BP would be 13 kPa at heart level; at brain level it would be – 8.1 kPa (13 kPa – $\rho g h_1$); at legs 25.74 kPa (13 kPa + $\rho g h_2$), where: h_1 – the distance between heart and brain levels (≈ 0.5 m); h_2 – distance between heart and lower limbs (≈1.3 m); ρ – blood density.

Blood pressure in adults in horizontal and vertical position is expressed in cm per column of blood and is represented in figure 6.8.

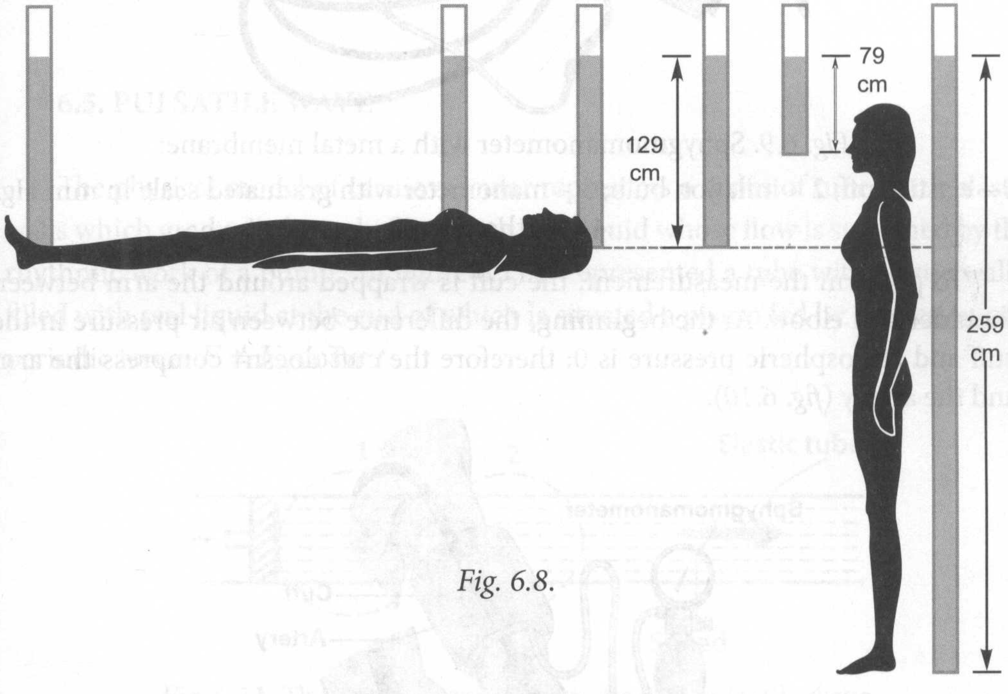


Fig. 6.8.

Korotkoff method to determine blood pressure

Blood pressure is very important as a physical parameter in diagnostics of various diseases. To determine either systolic or diastolic pressure in a certain artery, it is used a hollow needle connected to a manometer. In medical practice there's also used another method, which doesn't require penetration into the artery, suggested by N. Korotkoff. There are several modifications of this device, on of them, represented in figure 6.9, is called *sphygmomanometer*.

Fig. 6.9. Sphygmomanometer with a metal membrane:
1 – elastic cuff; 2 – inflation bulb; 3 – manometer with graduated scale in mm Hg; 4 – stethoscope.

To perform the measurement, the cuff is wrapped around the arm between shoulder and elbow. At the beginning, the difference between air pressure in the cuff and atmospheric pressure is 0; therefore the cuff doesn't compress the arm and the artery (*fig. 6.10*).

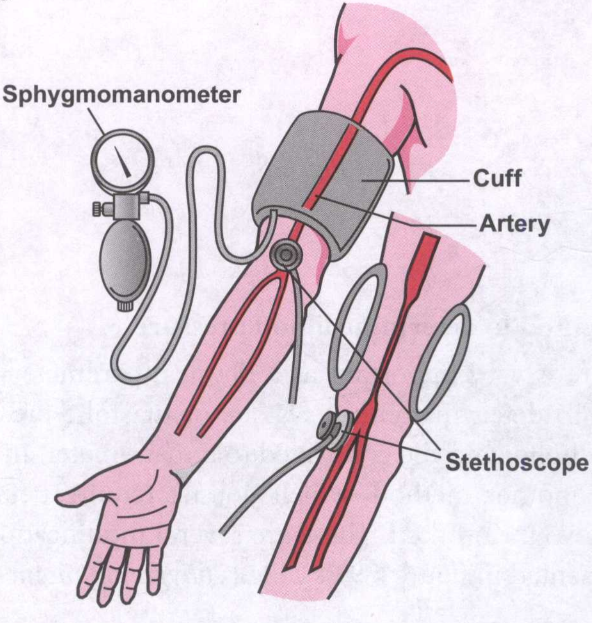


Fig. 6.10.

As we pump air in the cuff, the pressure excited by it upon the brachial artery grows and stops blood flow at one point.

This happens when air pressure from the cuff indicated by the manometer is bigger than the arterial pressure. Slowly opening the valve situated close to the inflation bulb we reach the moment when air pressure from the cuff is equal with **systolic pressure** – the blood starts passing through the narrow lumen of the artery, with **turbulent flow**.

While measuring the pressure, the doctor pleces the stethoscope on the artery distal to the cuff and auscultates the characteristic sounds which accompany this process.

6.5. PULSATILE WAVE

The physical model of cardiovascular represents a series of tubes with elastic walls which gradually branch, filled with real liquid whose flow is sustained by the rhythmic work of a pump. In figure 6.11 is represented a tube with elastic walls, filled with real liquid at the end of which is situated a piston led by the action of a periodic force: $F = F_0 \cos \omega t$.

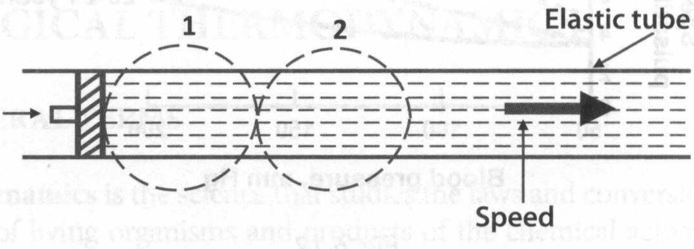


Fig. 6.11. The mechanism of formation of pulsatile wave.

Because of elastic walls the pressure created nearby the piston will move along the tube as pulsed wave.

The speed of this wave depends on several factors such as density of the liquid (ρ) and Elastic modulus (E) of the walls. We have a similar phenomenon in blood arteries.

During the phase when blood is ejected from the ventricle, there is produced a pressure wave which is spread in arteries with a certain speed (called **pulse speed**) which is much bigger than the speed of flowing blood. For example, in normal conditions, the average blood speed in aorta is 0.2–0.4 m/s, but the speed of pulse in aorta is approximately 4–6 m/s.

The faster the pressure wave spreads, the higher blood pressure will be. The speed of the pulse is bigger if the arteries walls are more rigid according to **Moens equation**:

$$V = F\sqrt{\frac{Ed}{2\rho r}} \qquad (4).$$

Where: ρ – is blood density; E – elastic modulus of the arterial walls; d – wall thickness; r – inner radius of the blood vessel; F – constant factor.

The elasticity of blood vessels of an older person decreases. Therefore, E, the speed of the pulse and blood pressure increase (a more elastic wall softens the pressure wave so the pulse speed in elastic vessels is smaller). In figure 6.12 is represented the variation between blood pressure and pulse speed depending on age.

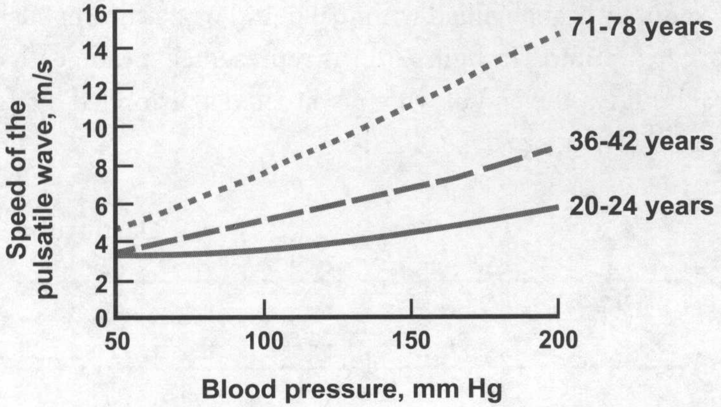


Fig. 6.12.

Determination of the motion speed of the pulsatile wave is used to diagnose cardiovascular diseases. Most of antihypertensive drugs decrease the arterial rigidity, behaving as antagonists (blocking factors) at the level of calcium channels in plasmatic membrane of muscular cells, situated in arterial walls which determines the relaxation of muscular fibers from the blood vessels walls. This is how blood vessel rigidity and vasoconstriction is prevented.

EXERCISES

1. The average speed of blood flow through aorta (v_a) is 30 cm/s, the speed of erythrocytes through capillaries (v_c) – is 0.5 mm/s. Based on these data, determine how many times the summed – area of all capillary sections is bigger than the area of a section from aorta.

Answer: of 600 times.

2. Determine the maximal quantity of blood which can go through an aorta in one second so its flow remains laminar. The diameter of aorta (d) is 2 cm, blood viscosity (η) – 5 $Pa \cdot s$.

Answer: 157 g.

3. In the avarage person, the aortic lumen has an area of 8 cm^2. The diameter of one capillary is 10^{-5} m. The blood flow speed decreases from 0.5 $m \cdot s^{-1}$ in aorta to 0.5 $mm \times s^{-1}$ in capillaries. Determine the total number of capillaries.

Answer: $10.2 \cdot 10^9$.

7. BIOLOGICAL THERMODYNAMICS

7.1. GENERAL TERMS

Thermodynamics is the science that studies the laws and conversion of energy. The existence of living organisms and products of the chemical activities in those organisms have a close interconnection with energy transformation, variation of equilibrium in the organism–environment system. Thermodynamics has two main principles, but before beginning to study them, we need to get acquainted to other terms that are commonly used in this science.

System – a set of interacting components, limited in an external environment. There are three types of systems classified according to their interaction with the environment. A system that has no matter or energy exchange with the environment is called **isolated**. A **closed system** only exchanges energy with the environment. An **open system** has energy and matter exchange with the environment.

Biological systems as well as the human body are open systems.

Energy – it is a quantitative property of movement, which transforms from one type to another. It characterizes the movement of matter, so the energy always determines the possibility of a system to do work. There are different types of energies.

Mechanical energy is the energy that describes the movement of objects and the power to do work when an object is moving. The mechanical energy is different from the **kinetic energy** – the energy of moving object and **potential energy** – the energy determined by the position of the objects.

Thermal energy is the sum of the **kinetic energies** of all the atoms and molecules of a substance resulting from their chaotic movement. Heat indicates the thermal movement. The value of the kinetic energy $\mathbf{E}_c$ for a monoatomic particle depends on the temperature **T**: $E_C = \frac{3}{2}KT$, where: K – Boltzmann's constant.

Chemical energy is the energy of interaction between the atoms and molecules. It represents the electrons movement on their outer electronic orbital of atoms and molecules.

Electric energy is the energy of interaction between the particles of the electrical field that leads to their movement through the electric field.

Work – it is the transformation of energy.

A thermodynamic state of a system is represented by the values of all the sets of parameters (measurable physical values) of the system at a certain moment of time.

There are two types of parameters:

– **Intensive** – they have different values in every point of the system and do not depend on the system size (pressure, concentration, temperature);

– **Extensive** – they depend on the system size and on the substance amount (volume, mass, number on moles).

The thermodynamic equilibrium has the following properties:

– State parameters are constant;

– Maximum disorder (maximum entropy);

– There are no more energy exchanges between the elements of the system and the environment;

– Entropy production stops.

The stationary state has the following properties:

– Stated parameters are constant;

– Intensive parameters are not constant in space;

– The exchange of matter and energy between the elements of the system and the environment do not stop;

– Entropy production has a minimal value, but it is not equal to 0.

Thermodynamic processes – the state transformation of a system with the change in stated parameters. Such processes can be:

– **reversible** –quasistatic processes; the system is always at equilibrium; the changes in some properties of the system will make the system to evolve from the final to the initial stated on the same path;

– **ireversible** – non-quasistatic processes, the system reaches its initial state (if it is possible) through another path and external force (it is not spontaneous);

– **cyclical** – the initial and final state of the system are identical; such processes can be irreversible.

7.2. FIRST LAW OF THERMODYNAMICS

After failing to build a machine that once activated would continue to function indefinitely without any energy consumption (perpetual motion machine of the first kind), it was concluded that energy cannot be "created", so to provide energy it is necessary to apply another one.

In 1840, Gh. Hess deduced that heat evolved or absorbed in a chemical reaction does not depend on the intermediate steps of the reaction, but only on the initial and final state of the reactants. Between 1842 and 1850, many scientists (J. R. Mayer, J. Joule, H. Helmholtz) discovered the equivalence between work and energy and determined the mechanical equivalent of calorie (**1 calorie = 4.18 joules**).

First law of thermodynamics or the principle of conservation of energy postulates the existence of a parameter that is typical for every system, called **internal energy** (U). This parameter expresses the capacity of a system to do work and has well defined value for every state. According to another statement from this law, the change in the internal energy (ΔU) is equal to the sum between the heat (Q) and work (mechanical, chemical, osmotic, electric, etc.) performed by the surroundings:

$$\Delta U = Q + L \qquad (1).$$

Or when the work is performed on the system and heat is received, in accordance we obtain:

$$\Delta U = Q - L \qquad (2).$$

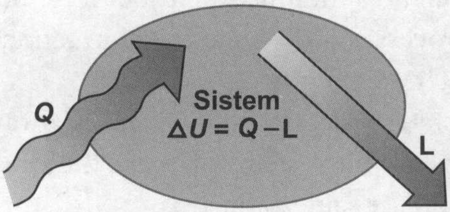


Fig. 7.1.

We are going to examine two different processes of a system and the surroundings in accordance to the first law, causing the same internal energy change (*fig. 7.2*).

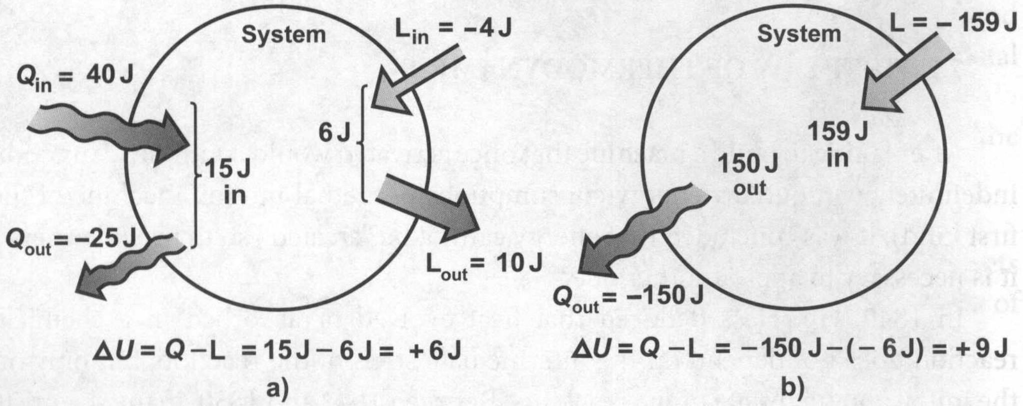


Fig. 7.2.

The internal energy of a system is the sum of its kinetic and potential energies (the sum does not include the kinetic and potential energies of the system as a whole due to its placement in the environment).

The heat (**Q**) represents the internal energy change of the molecules' random movement (thermal agitation), and **work** (**L**) – the internal energy modifications due to macroscopic movements (volume changes due to an external pressure, area changes due to surface tension, and transportation of electric charge in a potential difference etc.).

From the previous expressions we can deduce a third expression for the first law:

The energy of an isolated system is not created or wasted, but it just transforms from a form to another. In an isolated system $\Delta U = 0$.

The internal energy change can be practically measured through the heat change between the system and the environment, when it does not perform any work and the volume is constant (isochoric process). In cases of the biological processes that occur in the atmosphere, the pressure is therefore constant (isobaric process). In this case we introduce a new term that replaces the internal energy (U) – **enthalpy, (H)**.

If we write the equation for the first law of thermodynamics:

$$\Delta U = Q - L = Q - p\Delta V \quad (3),$$

heat (Q) is equal to:

$$Q = \Delta U + p\Delta V = \Delta(U + pV) = \Delta H \quad (4).$$

Therefore: $\Delta H = Q_{isobar}$ (temperature is changed by the isobar system).

$H = U + pV$ is called the **enthalpy** of a system and it is very useful for the thermodynamic studies of the chemical reactions. When $\Delta H > 0$, the system absorbs heat (**endothermic reactions**) and when $\Delta H < 0$, the system releases heat to the surroundings (**exothermic reactions**).

First law of thermodynamics in biology

Biological systems are open thermodynamic systems, and biological processes are irreversible thermodynamic reactions. The internal energy in living organisms can decrease or increase due to different conditions (age, physiological state etc.). For a correct application of the first law for organisms, we must take into account that these are open systems that absorb and release energy, so the conservation of energy is for a closed system made of organism and its surrounding.

7.3. ENERGY FLOW IN THE BIOSPHERE

The main energy source for the living beings is the light energy (*fig. 7.3*).

A small percentage of the light energy is absorbed by the plants for photosynthesis and used for converting from CO_2, H_2O and **minerals**, substances with a high energy.

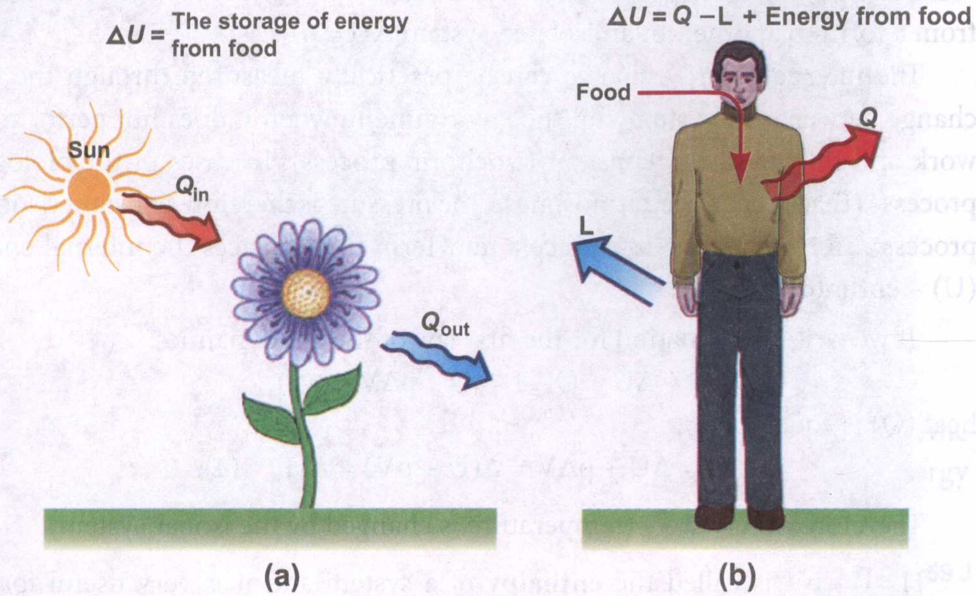


Fig. 7.3.

Chloroplasts are organelles in green plants which absorb the energy from the sunlight and form polysaccharides $(C_6H_{12}O_6)_n$ from H_2O and CO_2 and releases O_2. In **mitochondria**, (organelles from the cells of heterotrophic organisms) the products resulting from photosynthesis are used as a source of energy for **generating ATP**. Heterotrophic organisms are either parasites or live within living organisms. The ATP molecule is cleaved and the released energy is used to do mechanic, chemical, osmotic or any other type of work.

Energy balance of the organism

According to the first law of thermodynamics the can state that:

Energy intake = work produced + heat released + the stored energy of the organism.

The clinical test of basal metabolism intensity reports the effects of different factors on the metabolism energy. This is an example of balance in simplified conditions.

The studied object is at rest (it is not doing any work) and have not eaten in 12 hours (it is not absorbing any energy from the environment).The energy balance in this case is:

Heat released = – the stored energy = energy used.

7.4. THE SECOND LAW OF THERMODYNAMICS

The first law of thermodynamics proves us that energy can be transformed from one form to another and we can mathematically calculate these changes. However, the first law does not say anything about the direction of these transformations. We can find the answer to this question from the second law of thermodynamics. This principle was first founded by Carnot (1824) and later developed by Clausius (1850), Thomson (1851), Boltzmann (1880) etc. This principle sums up the practical impossibility of a thermic machine to spontaneously convert thermal energy into mechanical work (**perpetual motion machine of the second kind**) by doing a useful work.

According to the second law, there is a parameter **S**, called **entropy** that is a common function of state for every state of the system. **Entropy** measures the disorder of a thermodynamic system.

According to Clausis' macroscopic concept, if heat (**dQ**) is reversibly absorbed by a system at a temperature **T** (**isotherm**), then there is a function of state **S** which increases with **dS** as follows:

$$dS = \frac{dQ}{T} \quad (5).$$

In accordance with the definition of entropy, the second law of thermodynamics states that an isolated system cannot evolve to the state of order, mathematically:

$$dS \geq 0.$$

The equality sign reflects the situation when reversible transformations take place into an isolated system, and ">" reflects the situation when irreversible transformations happen in the system.

The process of which result is entropy decrease is called coupled process, and the process in which are increases of the entropy is called process cuplant. That coupled processes cannot take place on the account of coupled processes. For example, anabolism can only take place with catabolism, constituting the metabolism (*fig. 7.4*).

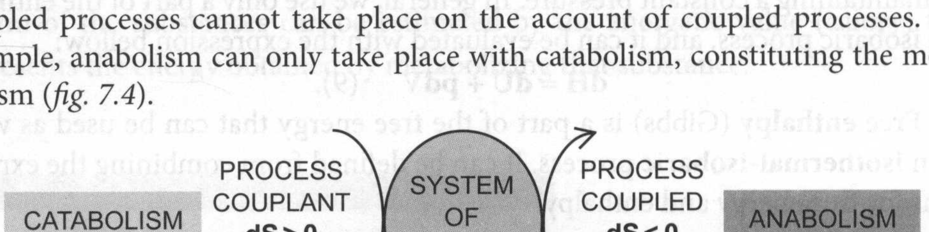


Fig. 7.4.

The thermodynamic equilibrium is a state with **maximum entropy**, and **stationary state** with **constant entropy**. The entropy produced by the irreversible processes of a system is released to the environment (a cell at rest is in stationary state). An equivalent statement for the second law of thermodynamics states that in a cyclical transformation, the received heat cannot be totally transformed into work. Although, the living systems are very energetically efficient, they cannot use the whole energy obtained (for example through cleaving the macro energetic compounds like ATP, etc.).

7.5. THERMODYNAMIC FUNCTIONS

Theoretically, we can measure the whole internal energy of a system through work. In reality, only a part of the internal energy can be used because of the heat loss. Therefore, the internal energy that can be used as mechanical work in an isothermal transformation (**T = const.**) is called **free energy** (**F**):

$$\mathbf{F = U - TS} \qquad (6).$$

The product **TS** represents the internal energy of the system that is lost as heat in an isothermal transformation. Usually, only a part of **F** is used in such transformations, and it can be calculated from the following expression:

$$\mathbf{dF = dU - TdS} \qquad (7).$$

As it was mentioned earlier, the internal energy of a system that can be used as work in an isobaric process (**P = const.**) is called enthalpy (**H**):

$$\mathbf{H = U + pV} \qquad (8).$$

In an isobaric process we can use the internal energy of the system to do work, but also to get some additional quantity of energy (**pV**), which is used by the system for maintaining a constant pressure. In general, we use only a part of the enthalpy in a isobaric process, and it can be evaluated with the expression bellow:

$$\mathbf{dH = dU + pdV} \qquad (9).$$

Free enthalpy (Gibbs) is a part of the free energy that can be used as work in an **isothermal-isobaric** process. It can be defined from combining the expressions for free energy and enthalpy:

$$\mathbf{G = U + pV - TS} \qquad (10).$$

In an isothermal-isobaric process we are going to use only a part of the available free enthalpy, as given in the equation:

$$\mathbf{dG = dU + pdV - TdS} \qquad (11).$$

The processes that occur in the human body are isothermal-isobaric and only the free enthalpy can be used from the internal energy (for example a compound macroergic). Also, only the free enthalpy from the internal energy of some products (food, energy products, etc.) can be efficiently used. The biological systems need to synthesize the complex molecules from the simple compounds for tissue renewal and growth, but also for creating the substances that have vital importance for the regulation of biological processes (hormones, enzymes, etc.). The needed energy provided for the functioning of the organism at rest as well as for other activities is an essential function of the biological systems.

Carbohydrates, **lipids**, **proteins**, **vitamins**, **minerals** and **water** are substances with great importance for our nutrition. The primary energy is obtained through catabolism, where the isobaric combustion of carbohydrates and proteins creates the free enthalpy. Enthalpy change through isobaric combustion of different biological substances (catabolism) is almost the same for amino acids and carbohydrates, but it is triple for lipids, which explains their function of storing energy in the organism. So, for **glycine** $\mathbf{\Delta H}$ = 3.12 kcal/g, for glucose $\mathbf{\Delta H}$ = 3.74 kcal/g, but for tripalmitin $\mathbf{\Delta H}$ = 9.3 kcal/g.

The primary energy cannot be directly used for maintaining all the vital functions, instead it is stored in the macroergic compounds so it can be released when necessary (for example: $ATP + H_2O \rightarrow ADP + P$). Basal metabolism is a cellular process essential for mantainning all the vital functions of the cells. It is obvious that this process will take place without any energy consumption. An important part of the energy consumption (approximately 66%) in this process is necessary for the ion pumps. The average yield of a metabolic process is approximately 20%. This shows that a big part of the energy is lost through heat. The energy contribution of the substances can be estimated by the energetic value (calory) that represents the energy obtained by metabolizing that substance.

EXERCISES

1. The calorific power of gasoline is $1.3 \cdot 10^8$ J/gal. The other factors, like temperature of the car are constant. What is the change of the internal energy of the car if there are 12.0 gals of gas?

Answer: $1.6 \cdot 10^9$ J.

2. A system performs work of $1.8 \cdot 10^8$ J, releasing $7.5 \cdot 10^8$ J of heat. Calculate the internal energy change of the system, knowing that there are no other changes in the system (like temperature or addition of fuel).

Answer: $-9.3 \cdot 10^8$ J.

3. A person did work of 500 J, releasing 9500 J of heat to the environment. Calculate: a) the internal energy change; b) the yield.

Answer: (a) $-1.0 \cdot 10^4$ J; (b) 5%.

4. (A) How much energy from the food will a person metabolize when doing work of 35 kJ, with a 5% yield? (B) How much isothermal heat was released?

Answer: (A) 700 kJ; (B) 665 kJ.

5. (A) Calculate the average metabolic rate (in watts) of a person that metabolizes 10 500 kJ of energy from food during 24 hours. (B) What is the maximal value of work that can be constantly done by a person by burning fat, considering that the maximal yield is 20%?

Answer: (A) 121 W; (B) $2.1 \cdot 10^6$ J.

6. A person uses 450 g of carbohydrates, 120 g of proteins, and 100 g of lipids a day for intellectual work. Calculate the energy produced after the combustion of these substances. Consider that only 90% of the substances assimilate.

Answer: 12.2 MJ.

7. Through physical work, a person receives 17 MJ from food. He does work of 10 MJ during the day. What part of the received energy is transformed into useful work?

Answer: 0.59.

8. What is the value of the useful work that can be done by combusting a mole of glucose, considering that the human body works as a thermal machine with a 30% yield.

Answer: 854 kJ.

8. MOLECULAR TRANSPORT PHENOMENA

8.1. DIFFUSION. FICK'S LAW

Diffusion represents the transport of a substance between two areas with different concentrations or electric potential due to thermal agitation.

We are going to use a measuring cylinder that contains a solution of **eosin** (a red colored substance) at the bottom and water on top (*fig. 8.1*). We can see that the eosin molecules or ions spread out through the water solution. This phenomenon represents a **simple diffusion**, because it occurs without an external force.

In terms of quantity, **Fick's laws** describe the diffusion in a homogenous environment.

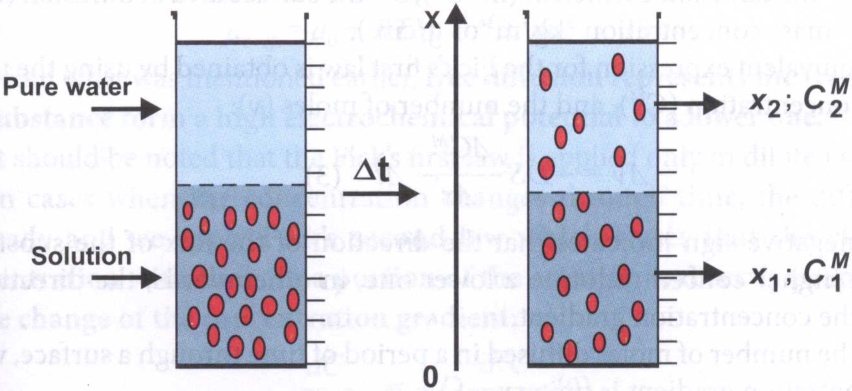


Fig. 8.1.

Before starting to analyze the diffusion laws, we need to get familiar with some terms such as **diffusion flux** and **concentration gradient.**

The diffusion flux represents the amount of substance, energy etc. that *pass* per unit area and per unit time.

In diffusion phenomena are most commonly used:

$$\text{the mass flow rate, } \frac{\Delta m}{\Delta t}\text{, molar flow rate, } \frac{\Delta \nu}{\Delta t}.$$

Where: m – mass of the substance; ν – the number of moles.

So, in a simpler way we can say that **the gradient is the change of a measure (concentration, density, electric potential, etc.) between two points, depending on the distance between the points.**

For diffusion we use either density gradients ($\frac{\Delta\rho}{\Delta x}$), or concentration gradients ($\frac{\Delta c}{\Delta x}$).

Fick's first law postulates that the mass of the normally diffused substance through a surface is directly proportional to the surface area, concentration gradient and time.

For example researched (*fig. 8.1*) that the weight concentration gradient is determined from the relationship:

$$\frac{\Delta c^M}{\Delta x} = \frac{c_2^M - c_1^M}{x_2 - x_1} \quad (1).$$

Thus, Fick's law has the following mathematical expression:

$$\Delta m = -DS\frac{\Delta c^M}{\Delta x}\Delta t \quad (2).$$

Where: D – the diffusion coefficient ($m^2 \cdot s^{-1}$); S – the surface area of diffusion (m^2 or cm^2); c^M – mass concentration (kg/m^3 or g/cm^3).

An equivalent expression for the Fick's first law is obtained by using the terms of mass concentration (C^M), and the number of moles (ν):

$$\Delta\nu = -DS\frac{\Delta C^M}{\Delta x}\Delta t \quad (3).$$

The negative sign indicates that the direction of the flux of the substance is from a higher concentration to a lower one, in other words, the direction is towards the concentration gradient.

D is the number of moles diffused in a period of time through a surface, when the concentration gradient is unitary.

In a fluid environment, the diffusion coefficient is determined by **Einstein's equation:**

$$D = \frac{KT}{f} \quad (4).$$

Where: T – absolute temperature; K – Boltzmann's constant ($1.38 \cdot 10^{-23}\ J \cdot {}^{\circ}K^{-1}$); f – friction coefficient of the moving particles in a fluid environment.

For the spherical molecules (particles) with radius ***r*** that move through a fluid with a viscosity **η** and constant velocity the friction force is given by **Stokes' law**:

$$F = 6\pi\eta r v.$$

In this case, the friction coefficient is equal to $6\pi\eta r$ and the diffusion coefficient to:

$$D = \frac{KT}{6\pi\eta r} \quad (5).$$

Electrochemical potential

Depending on its concentration, the solute has a certain potential energy, called **chemical potential – μ_K**. It is common for every substance from the solution and it is determined by the following expression:

$$\mu_{Ki} = \mu_{oi} + RT\ \ln C_i^M \quad (6).$$

Where: i – solute's component; C_i^M – molarity; R – the gas constant; T – absolute temperature; μ_{oi} – standard potential (unitary molarities solvent's chemical potential at 25°C).

There is electrical potential energy (μ_e) in an electrolytic solution, which for a mole of ions is determined by the expression:

$$\mu_e = V\,n\,F \quad (7).$$

Where: V – electric potential; n – the valence of the electron; F – Faraday's number.

The total potential energy is called the **electrochemical potential (μ_{e-K})** and it is determined by the following equation:

$$\mu_{e-K} = \mu_0 + RT \ln C^M \pm VnF \quad (8).$$

From what was mentioned earlier, **free diffusion represents the transport of the substance form a high electrochemical potential to a lower one.**

It should be noted that the Fick's first law is applied only in diluted solutions. In cases when the concentration changes through time, the diffusion is unsteady and we apply **Fick's second law which states that the change of concentration with time, in a portion of the solution is directly proportional to the change of the concentration gradient:**

$$\frac{dC}{dt} = -D\frac{d^2C}{dx^2} \quad (9).$$

The diffusion coefficient (D) (the same from Fick's first law) is directly proportional to the diffusion velocity of the particles, which depends on temperature, viscosity of the solution and size of the particles.

8.2. DIFFUSION THROUGH ARTIFICIAL MEMBRANES

All the known types of membranes are divided into two big groups – **artificial** and **natural (biological)**. Artificial membranes can be **semipermeable** and **selectively permeable**. The semipermeable membranes allow the molecules of the solvent to pass through it but do not allow the molecules of the solute to pass through.

As a consequence, the solvent diffuses from the hypotonic solution into the compartment with hypertonic solution. Such diffusion is called **osmosis** (*fig. 8.2*).

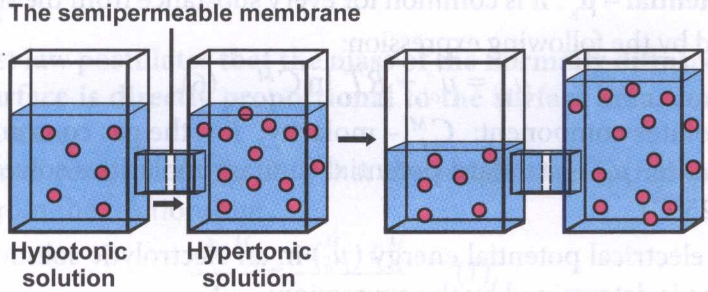


Fig. 8.2.

Artificial membranes like the ones from plastic or cellophane are of great importance in scientific researches or in medical practice.

A common property of all these types of materials is that they are porous substances that allow micro molecules to pass through them, but not the macromolecules. For example, artificial membranes are used for **dialysis** in medical practice. Dialysis is the principle of functioning for the **artificial kidney** and allows the purification of patient's blood with **renal failure** once every 2 or 3 days.

The device (*fig. 8.3*) that is used represents a system with capillary membrane tubes that allows the blood to circulate throughout the procedure. In an opposite direction circulates the dialysis liquid.

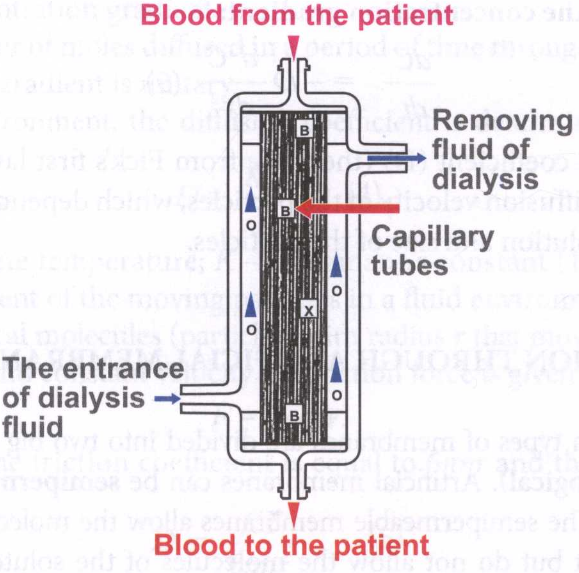


Fig. 8.3.

The scheme with the principle of such an installation is shown in the (*fig. 8.4*). The equivalent membranes to the dialysis membranes can be used for purifying the solutions that contain particles of different sizes in pharmaceutical and physicochemical researches.

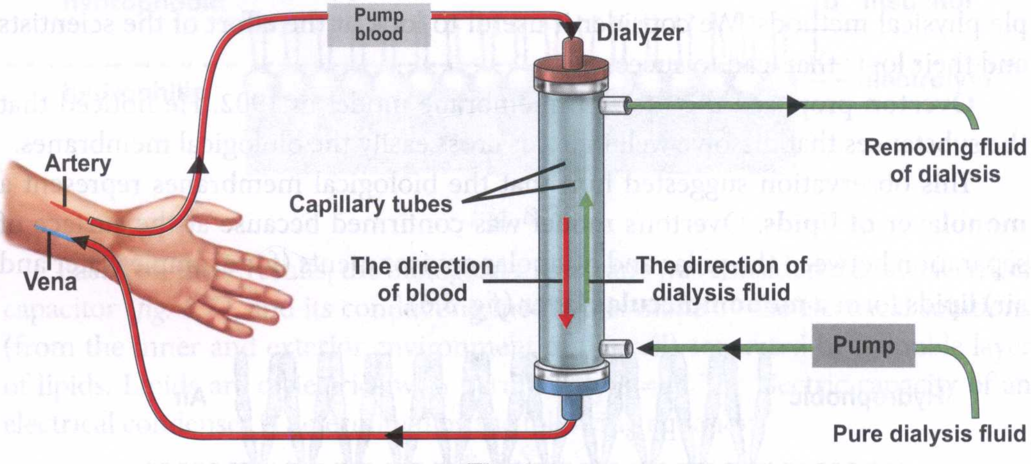


Fig. 8.4.

8.3. DIFFUSION THROUGH CELL MEMBRANES. PASSIVE TRANSPORT

Physical methods to study the composition and structure of cell membranes

Cell membrane is a hydrophobic structure, which divides intra- and extracellular aqueous solutions.

Different biochemical and biophysical methods of examination determined that the composition of all cell membranes consist of: proteins, lipids, carbohydrates, glycoproteins, lipoproteins, ions and water.

Proteins are the main macromolecules responsible for the spatial organization of cell membranes and their functioning.

Numerically, lipids have the highest proportion in membrane's structure. The most important lipids are phospholipids and cholesterol.

Lipid molecules consist of a hydrophilic ending and two hydrophobic chains (*fig. 8.5*).

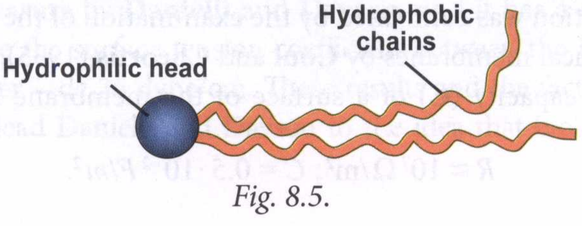


Fig. 8.5.

Throughout time, the spatial arrangement of the components of the cell membrane was a very difficult problem.

It was possible to make an image of the structure and molecular composition of the cell membranes before the electronic microscopy with some relatively simple physical methods. We consider it useful to look at the effort of the scientists and their logic that lead to success.

Overton proposed the first cell membrane model in 1902. He noticed that the substances that dissolve well in lipids cross easily the biological membranes.

This observation suggested him that the biological membranes represent a **monolayer of lipids**. Overton's model was confirmed because at the surface of separation between the polar and nonpolar environments (for example water and air) lipids form a **monomolecular layer** (*fig. 8.6*).

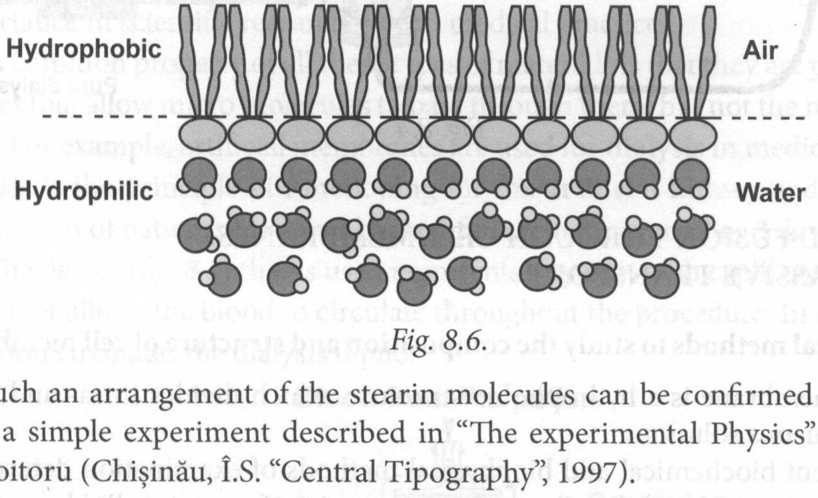


Fig. 8.6.

Such an arrangement of the stearin molecules can be confirmed by performing a simple experiment described in "The experimental Physics" signed by D. Croitoru (Chișinău, Î.S. "Central Tipography", 1997).

In 1925, **E. Gorter** and **F. Grendel**, extracted the membrane lipids from a known number of erythrocytes using acetone. Conditions for the spreading into a monolayer were created for the extracted lipids. It was noticed that the surface area of the monolayer was twice the number of the sum of the surface areas of the erythrocyte membranes.

Based on the result of this experiment, they proposed a new model, where the cell membrane has a bilayer of lipids (*fig. 8.7*).

This assumption was confirmed by the examination of the electrical parameters of the biological membranes by **Coul** and **Cheortis** (1935). Electric resistance (R) and electric capacity (C) of a surface of the membrane have relatively big values:

$$R \approx 10^7\,\Omega/\mathrm{m}^2;\ C = 0.5 \cdot 10^{-2}\,F/m^2.$$

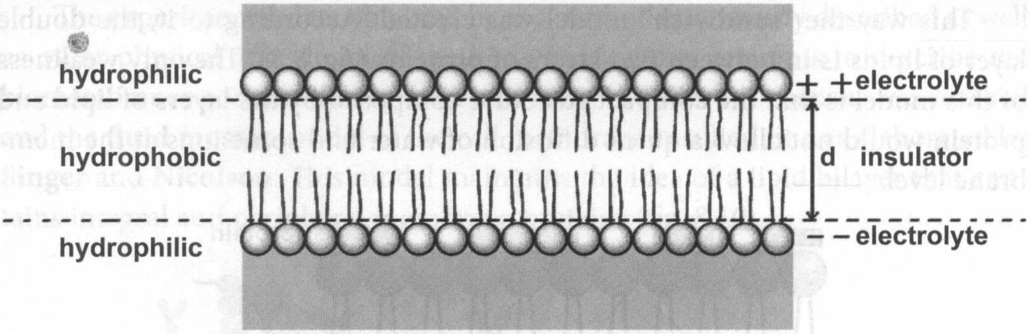


Fig. 8.7.

Based on these results, the biological membrane was considered as an electrical capacitor (*fig. 8.7*), and its conducting medium is made by the electrical solutions (from the inner and exterior environment of the cell) separated by a double layer of lipids. Lipids are dielectrics with permittivity $\varepsilon = 2$. The electric capacity of an electrical condenser is determined by the following equation:

$$C=\frac{\varepsilon_0\,\varepsilon\,S}{d} \quad (10).$$

Where: ε_0 – vacuum permittivity ($\varepsilon_0 \approx 8.85\cdot10^{-12}$ F/m); d – the distance between the capacitor plates.

The capacity exerted on a surface:

$$C=\frac{\varepsilon_0\,\varepsilon}{d} \quad (11).$$

From equation (11) it can be determined the distance between the plates, and it corresponds to the thickness of the lipid membrane:

$$d=\frac{\varepsilon_0\,\varepsilon}{d}\approx\frac{8.85\cdot10^{-12}\cdot 2\mathrm{m}}{0.5\cdot10^{-2}}=3.6\,nm.$$

It was confirmed that the obtained value is equal to the thickness of the nonpolar double lipid layer.

We mention that the hypothesis of a double lipid layer is maintained in all the other membrane models proposed.

The surface tension coefficient at the interface between the membrane and water was measure by **Danielli** and **Davson**, and it has a constant value of 0.2 dyne/cm, while the surface tension coefficient between the interface of lipids and water is higher – de 36 dyne/cm. These results and the fact that proteins are tensioactive have lead Danielli and Davson to the idea that biological membrane contains proteins.

This way the **"sandwich"** model was created. According to it, the double layer of lipids is in between two layers of proteins (*fig. 8.8*). The only weakness of this model is that the compact structure composed by the layers of lipid and protein would not allow a quick diffusion of water and some ions at the membrane level.

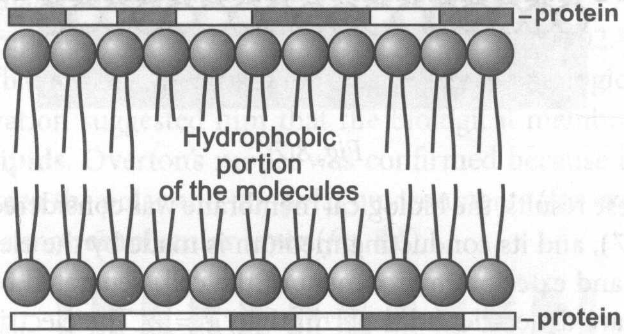


Fig. 8.8.

In 1956 the authors of this model admit the existence of membrane pores. The results obtained throughout the experiments emerged the necessity of rejecting the "sandwich" model. The physical methods of research (**Diffraction with X rays, electronic microscopy, cryodecapage**) performed with devices with a high precision are of great importance in the study of biological membrane structure.

The orderly arrangement of the lipid molecules in the double layer was confirmed by the X ray method and the sizes of the biomembrane structure was determined with different accuracy.

Electron microscopy was first applied by **Robertson** (1957). After the membrane was processed in advance, the electrogram showed three layers (*fig. 8.9*). The thickness of the membranes of different cells varies from 7 to 15 nm.

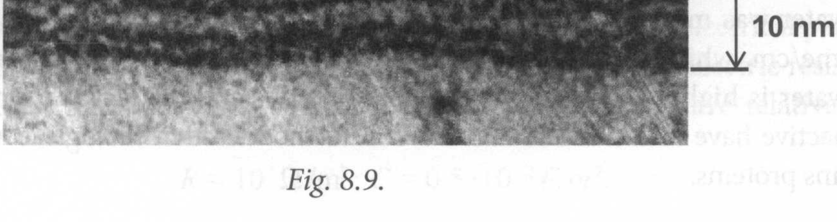


Fig. 8.9.

The experimental data obtained by the methods previously described as well as other physical and chemical methods, made other attempts to describe the biomembranes possible during 1960 and 1970. In 1972, a consensus was reached and the **fluid mosaic model** of the cell membrane resulted; it was elaborated by **Singer** and **Nicolson**. This model maintains the idea of a lipid bilayer that contains integral and peripheral membrane proteins (*fig. 8.10*).

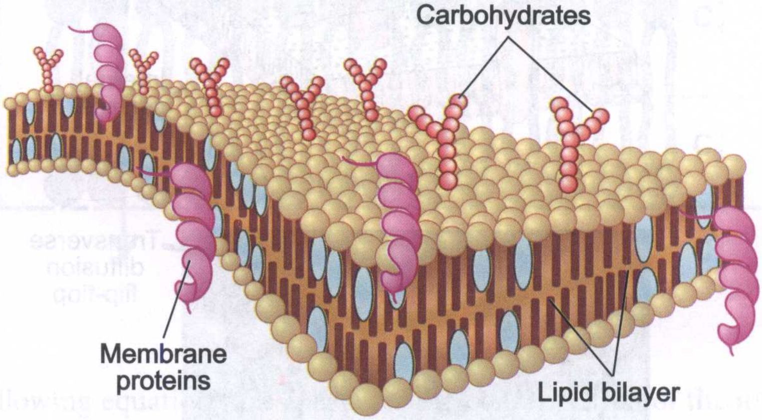


Fig. 8.10.

The fluid mosaic model was confirmed by many biophysical techniques. The most productive one is **cryofracture (cryodecapage)** of the cell membranes and then its examination at the **electric microscope**.

This method involves a few successive stages:
- accelerated cooling of the cell in liquid nitrogen;
- detachment of the cell membrane;
- its placement into a vacuum container;
- the membrane is cut longwise the lipid layers with a special knife;
- the obtained preparation is sprayed with platinum and carbon atoms;
- then it is investigated at the microscope.

The membrane is detached into two layers emphasizing the internal structure (*fig. 8.11*). The presence of integral proteins in the cell membrane was discovered through such an experiment.

The lipid bilayer is a dynamic structure presenting fluidity: the lipid molecules have a translational motion across their layer (**lateral diffusion**), conical rotation around their axis, flexion, and movement from a lipid layer to another (**transverse diffusion**) (*fig. 8.12*).

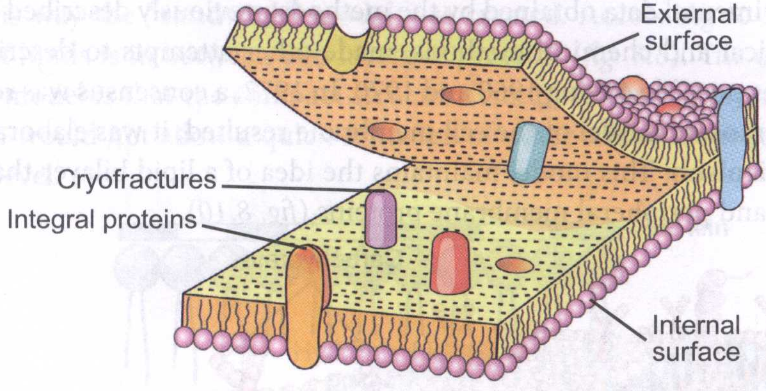


Fig. 8.11.

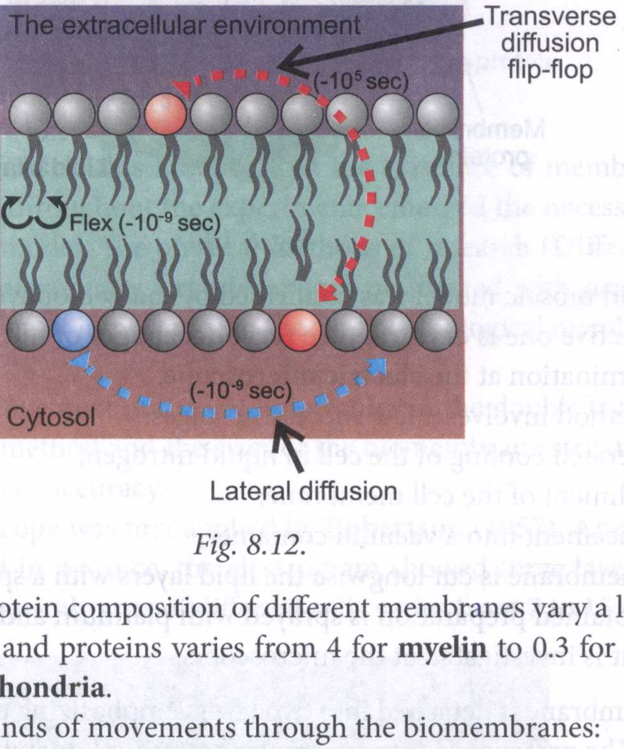


Fig. 8.12.

The lipid and protein composition of different membranes vary a lot; the mass ratio between lipids and proteins varies from 4 for **myelin** to 0.3 for the internal membrane of **mitochondria**.

There are two kinds of movements through the biomembranes:

1) **passive transport** – towards the electrochemical potential, without consumption of metabolic energy;

2) **active transport** – against the electrochemical potential with metabolic energy consumption.

Simple passive transport

For defining the term of permeability we are going to consider a biological membrane with a thickness of about 10 nm that separates aqueous solutions with the concentrations C_1 and C_2 ($C_1 > C_2$) (*fig. 8.13*).

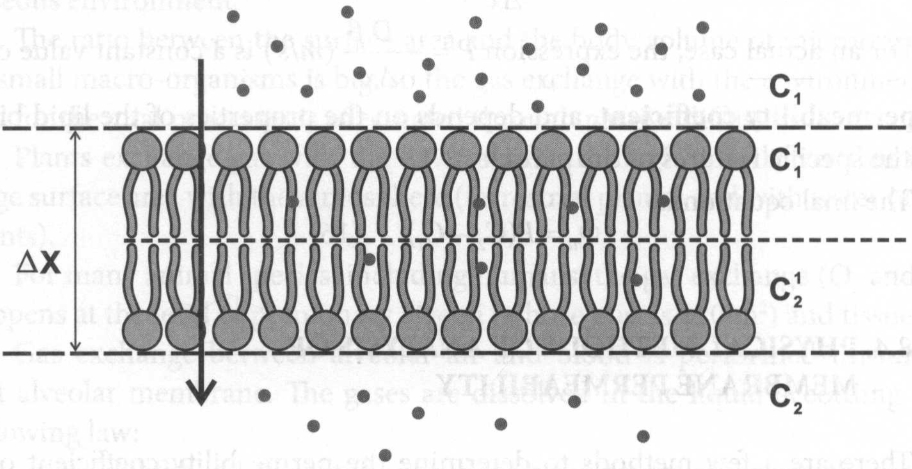


Fig. 8.13.

The following equation takes place for the concentration of the internal and external parts of the membrane:

$$\beta = \frac{C_1^1}{C_1} = \frac{C_2^1}{C_2} = \frac{C_2^1 - C_1^1}{C_2 - C_1} \quad (12).$$

Where: β – is called *partition coefficient*.

In such conditions the concentration gradient for the internal part of the membrane is:

$$\frac{C_2^1 - C_1^1}{\Delta x}.$$

The unitary flow of the substance (**Φ_u**) through the membrane is equal to the ratio between the mass of the solute (**Δm**) and the product between time (**Δt**) and the surface of the membrane (**S**):

$$\Phi_u = \frac{\Delta m}{S\Delta t} \quad (13).$$

By substituting **Δm** from the equation (2) in the equation (13), we obtain:

$$\Phi_u = -\frac{DS\Delta C\Delta t}{S\Delta t\,\Delta x} \Rightarrow -\frac{D\Delta C}{\Delta x} \Rightarrow -D\frac{C_2^1 - C_1^1}{\Delta x} \quad (14).$$

It is difficult to measure the concentrations from the internal part of the membrane, so for the equation (14) they can be substituted with the concentrations from the external part of the membrane, using the partition coefficient:

$$C_2^1 - C_1^1 = \beta(C_2 - C_1).$$

The equation for the **unitary flow** can be written as:

$$\Phi_u = -D\frac{\beta}{\Delta x}(C_2 - C_1) \quad (15).$$

For an actual case, the expression $P = \frac{D\beta}{\Delta x}(m/s)$ is a constant value called the **permeability coefficient**, and depends on the properties of the lipid bilayer and the specific features of the particle.

The final equation is:

$$\Phi_u = P(C_2 - C_1) \quad (16).$$

8.4. PHYSICAL METHODS OF ANALYSIS OF MEMBRANE PERMEABILITY

There are a few methods to determine the permeability coefficient of the membranes.

- The **osmotic method** consists of emphasizing the velocity of the volume change of the cells can be either hypertonic or hypotonic solutions.
- The **indicators method**, which is based upon the change in the color of the intracellular content when particular substances penetrate the cell.
- The **method of electrical conductivity** of the membrane. In certain conditions, when the measurements are taken at low frequencies of the alternating current, the electrical conductivity measures the permeability of the membrane. This method is used only for measuring the membrane permeability for ions.
- The **labeled atom method** is based upon using radioactive isotopes. The penetration or the way out of a substance in a cell can be detected with the contour of ionizing radiation.

The examination of the cell membrane permeability is of importance in medicine, especially for pharmacology and toxicology. The efficiency of the pharmaceutical products depends on the permeability of a membrane to those particular substances.

8.5. THE IMPORTANCE OF THE DIFFUSION OF GASES IN BIOLOGY AND MEDICINE

The diffusion of gasses (especially $\mathbf{O_2}$, $\mathbf{CO_2}$ and $\mathbf{N_2}$) occurs in the processes of respiration of plants and animals as well as photosynthesis.

Because the flux of the diffused substance is directly proportional to the surface area of diffusion, the aerobic organisms need a larger surface area with the gaseous environment.

The ratio between the surface area and the body volume of microorganisms or small macro-organisms is big, so the gas exchange with the environment needed for the vital processes is effectuated through passive diffusion.

Plants exchange gas with the environment through their leaves that have a large surface area with the atmosphere (terrestrial plants) and with water (aquatic plants).

For many animal species, including humans, the gas exchange (O_2 and CO_2) happens at the level of pulmonary alveoli (whose area is 120 m^2) and tissues.

Gas exchange between alveolar air and blood is performed through the wet alveolar membrane. The gases are dissolved in the liquid according to the following law:

– the solubility coefficient (**S**) depends on the nature of the gas, the nature of the solvent and the temperature:

$$S = \frac{\text{Maximum volume of gas solvit}}{\text{Volume of solvent}};$$

– for the given gas-solvent solution, at a constant temperature, the quantity of gas dissolved is directly proportional to the pressure of the gas exerted on the surface of the liquid;

– in the case of a gas mixture, every gas independently dissolves in the solvent. The quantity of every solute gas is directly proportional to the partial pressure of the mixture.

In the pulmonary alveoli, the gases are dissolved in the liquid, after that, they diffuse through the alveoli membranes into the blood vessels. These gases either remain dissolved in the blood (for example nitrogen) or can partially make labile (**oxyhemoglobin** and **carbohemoglobin**) or stabile chemical mixtures (**carboxyhemoglobin**).

It is important that the labile mixtures of O_2 and CO_2 with hemoglobin are formed and destroyed based on the partial pressure of these gases.

In alveoli, the partial pressure of oxygen is relatively high (100 mm Hg), which favors its passing through the venous blood (first dissolves and then diffuses). Here O_2 forms the oxyhemoglobin with hemoglobin.

At the tissue level, partial pressure of oxygen is lower (20–40 mm Hg) which implies its realease from the oxyhemoglobin. The partial pressure of CO_2 is higher (53–76 mm Hg) at this level, favoring the formation of carbohemoglobin.

At the alveoli level, the partial pressure of CO_2 is lower (40 mm Hg), and it leads to the release of CO_2 from the carbohemoglobin mixture.

In the conditions mentioned before, diffusion coefficinet for CO_2 is higher that the diffusion coefficient of O_2:

$$\frac{D_{CO_2}}{D_{O_2}} = 25.$$

And it is really important for the biological systems.

CO_2 is the most important product of catabolism in animals and it is easily eliminated from the organism and used as row material for photosynthesis and it is easily absorbed (through diffusion) in the atmosphere.

8.6. FACILITATED DIFFUSION

A lot of nutrients needed for the cell, being relatively large hydrophilic molecules, cannot cross through the membrane by simple diffusion. In such cases, the transport process is accomplished by the transmembrane channels, mechanism called **facilitated diffusion.**

The transmembranar channels recognize specific molecules (and sometimes ions) and exploit them under the activity of electrochemical gradient.

There are many methods of facilitated transport, one of them, called the conformational change, shown in figure 8.14.

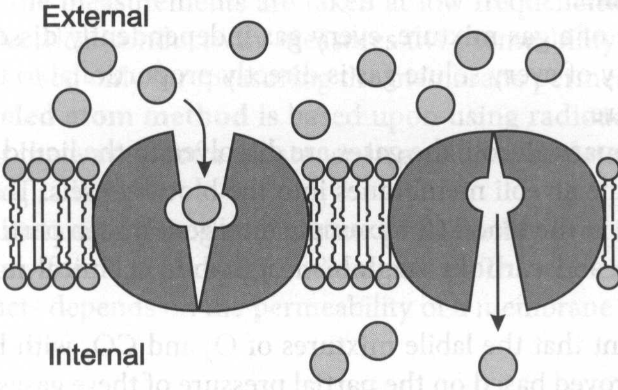


Fig. 8.14.

In comparison to simple diffusion, facilitated diffusion has the following features:

– lower efficiency than the one through the channels, but faster;

– saturation property: with the concentration increase of the substance, the flux of the substance increases only to a certain level, when all the transmembrane molecules are already occupied (*fig. 8.15*);

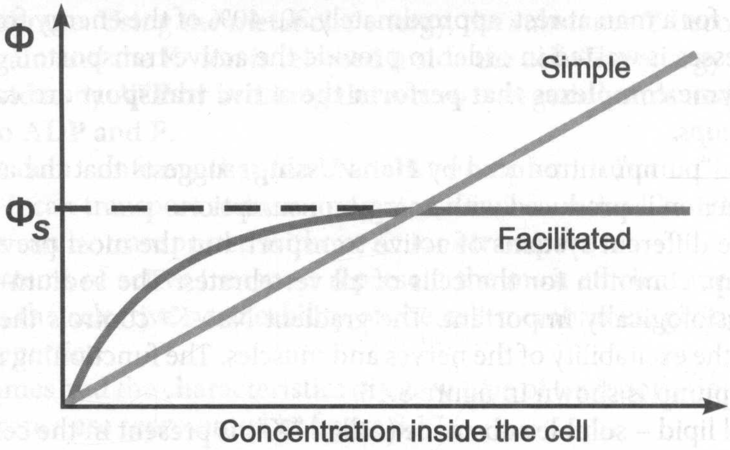


Fig. 8.15.

– competition of the transmembrane molecules is noticed: the substance with a higher permeability coefficient restrains the transportation of the substance with a lower permeability coefficient;

– there are substances that form stable compounds with the transmembrane molecule, blocking, in this way, the facilitated diffusion. For example, **flurazine**, which inhibits the transport of saccharides through the biological membrane. If the transport of a substance can be blocked, it can be concluded that it is being transported through facilitated diffusion.

8.7. ACTIVE TRANSPORT

The passive transport always has the tendency to homogenize the uneven distribution of substances in the intercellular and extracellular environments. Nevertheless, the cell composition is different from the composition of the extracellular environment. There are high concentrations of potassium ions, non diffusible ions, phospholipids, anions of amino acids and other ions of organic compounds in the cell, that are in extremely low concentrations in the extracellular fluids. On the contrary, there are other substances that are in higher concentrations in the external environment that they are in the cell (for example sodium ions). Those mentioned earlier and other experimental data show that together with the passive transport takes place the active transport of the substances as well.

The transport of molecules and ions against the electrochemical gradient, accomplished with the use of metabolic energy of the cell is called *active transport*.

For example, for a man at rest, approximately 30–40% of the energy from the metabolic processes is wasted in order to provide the active transport.

The enzyme complexes that perform the active transport are called *active transport pumps.*

The term "pump", introduced by **Hans Ussing**, suggests that the active transport phenomenon is produced with energy consumption.

There are different systems of active transport, but the most prevalent is the Na^+–K^+ pump, common for the cells of all vertebrates. The sodium–potassium pump is physiologically important. The gradient Na^+–K^+ controls the volume of the cells and the excitability of the nerves and muscles. The functioning mechanism of a Na^+– K^+ pump is shown in figure 8.16.

A special lipid – soluble substance (called "**X**") is present in the cell membrane, it is formed in the internal side of the membrane and it has special affinity to **Na^+** ion. It combines with a Na^+ ion from the intracellular environment forming the **X Na^+** compound.

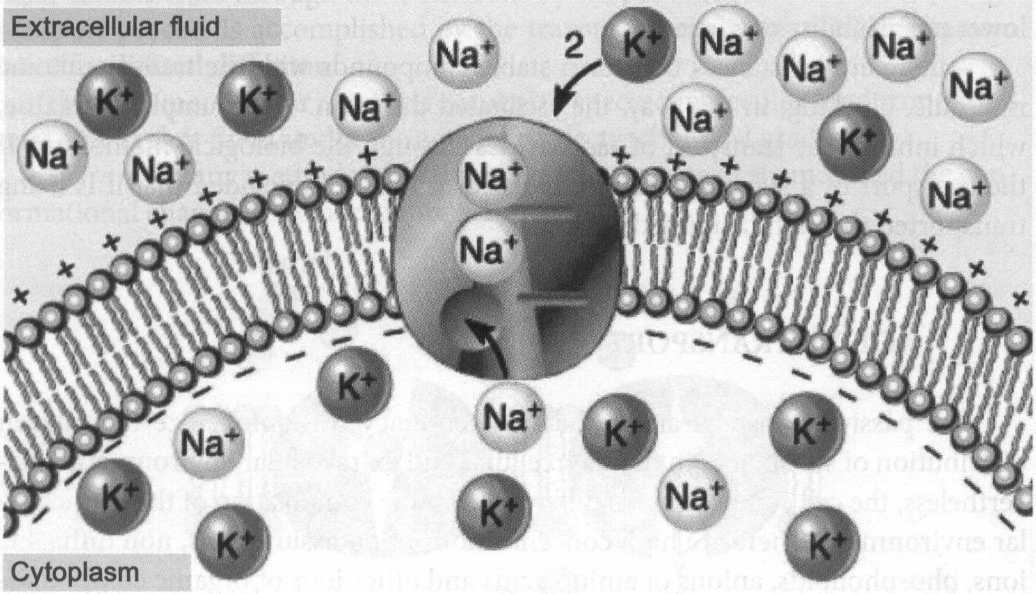


Fig. 8.16.

After the formation of this compound, a concentration gradient is produced between the two membranes, which determine the movement of XNa^+ to the external side of the membrane, where an enzyme transforms the substance X into substance Y. Y does not have affinity to Na^+ ion, so it is expelled into the extracellular environment. The Y substance has affinity to the K^+ ion from the extracellular environment forming the compound that moves to the external side of the membrane. Using the metabolic energy, the substance Y becomes the substance X again and the K^+ ion is released inside the cell. The energy for this reaction is provided by **ATP** by breaking the macro energetic bonds and the cleavage of ATP into **ADP** and **P**.

Recent data confirms that the **Na^+– K^+** pump can function in a couple of sequences. It can transport sodium and potassium ions ratios like **1:1, 1:2, 1:3, 2:3** or Na^+ ions can be transported without K^+ ions transport.

The systems of active transport represent the most efficient mechanism that determines the selective permeability of the cell membranes, giving them some specific properties.

The names and the characteristics of every type of transportation through the cell membranes are represented in figure 8.17.

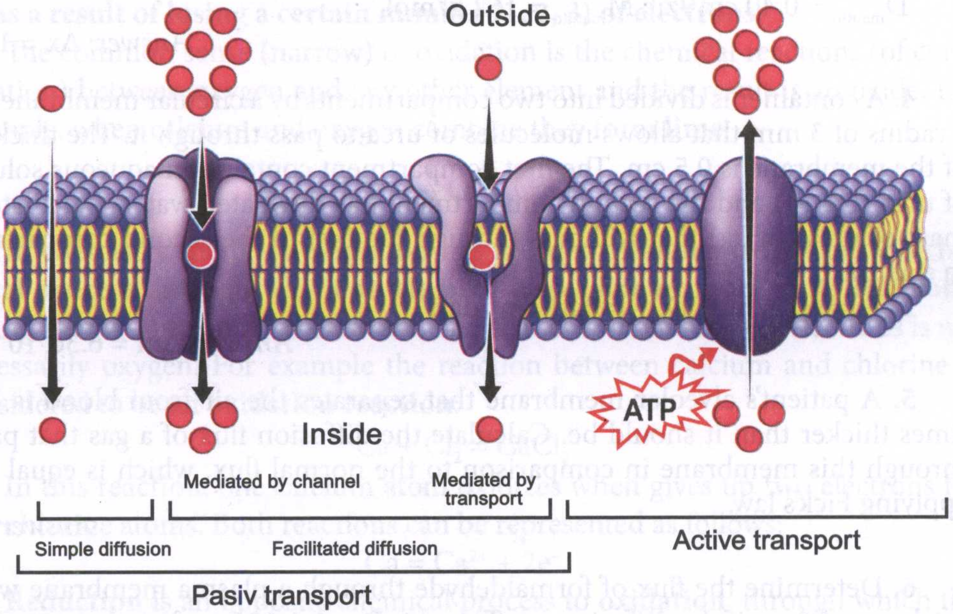


Fig. 8.17.

EXERCISES

1. A solution of hemoglobin with concentration 10^{-4} mol/l diffuses through a membrane with surface area of 20 cm^2, to a solution with a concentration of $1.4 \cdot 10^{-5}$ mol/l. The diffusion coefficient of hemoglobin is $D = 6.9 \cdot 10^{-7}\ cm^2 s^{-1}$; $M = 68\,000\ g \cdot mol^{-1}$. Find the mass of hemoglobin that passed through a distance of 5 cm during a minute.

Answer: $\Delta m = 9.6 \cdot 10^{-7}$ g.

2. With approximation, determine in Å, the radius of a spherical gas molecule at a temperature of 41°C.

There are given: – Viscosity coefficient $\eta = 10^{-5}\ \mathrm{Nm^{-2}s}$;
– Boltzmann's constant $K = 1.38 \cdot 10^{-23}\ \mathrm{j \cdot {}^{\circ}K^{-1}}$;
– Diffusion coefficient $D = 0.069 \cdot 10^{-6}\ \mathrm{m^2 s^{-1}}$.

Answer: $r \approx 3.3$ Å.

3. A container is divided into two compartments by a membrane with a surface area of 10 $\mathrm{cm^2}$, permeable for **urea** molecules as well as for **manito** ones. The concentration of urea is 24 g/l in both compartments and concentration of manito is 0.5 mol/l in the first compartment and it is equal to 0 in the second one. Knowing that the mass of manito that passes through the membrane in an hour is $1.5 \cdot 10^{-2}$ g. Calculate, in centimeters, the thickness of the membrane Δx.

$D_{manito} = 0.40\ \mathrm{cm^2/zi}$; $M_{manito} = 162$ g/mol.

Answer: $\Delta x = 1$ cm.

4. A container is divided into two compartments by a circular membrane with a radius of 3 mm that allows molecules of urea to pass through it. The thickness of the membrane is 0.5 cm. The first compartment contains an aqueous solution of urea (0.2 M) and the second compartment has distilled water. Calculate the mass of urea (in μg) that will pass through the first to the second compartment in a second.

$M_{urea} = 60$ g/mol; $D_{urea} = 0.81\ \mathrm{cm^2/day}$.

Answer: $\Delta m = 6.36 \cdot 10^{-2}$ μg.

5. A patient's alveolar membrane that separates the air from blood is 1, 25 times thicker than it should be. Calculate the diffusion flux of a gas that passes through this membrane in comparison to the normal flux, which is equal to 1, applying Fick's law.

Answer: 0.8.

6. Determine the flux of formaldehyde through a plasma membrane with a thickness of 8nm if the diffusion coefficient is $D = 1.4 \cdot 10^{-8}\ \mathrm{cm^2 s^{-1}}$, the concentration outside the cell at the initial moment is $C_0^M = 20 \cdot 10^{-4}$ mol/l and inside the cell is 10 times lower.

Answer: $\Phi \approx 3.2 \cdot 10^{-6}\ \mathrm{mol \cdot cm \cdot l^{-1} \cdot s^{-1}}$.

7. A membrane made of a bilipid layer with a thickness of 10 nm divides a container into two compartments. In one of the compartments the solution of **methyl blue** has a concentration of 10^{-2} mol/l and in the other one it is $2 \cdot 10^{-3}$ mol/l. Considering that the flux of methyl blue through the membrane is constant and equal to $3 \cdot 10^{-4}\ \mathrm{mol \cdot cm \cdot l^{-1} \cdot s^{-1}}$, determine the diffusion coefficient of this substance.

Answer: $D = 2.4\ \mathrm{cm^2 \cdot s^{-1}}$.

9. ELECTRODE POTENTIALS

Electrode potential is a fundamental concept of the electrochemical and corrosion theory, which helps predicting the direction and intensity of an electrochemical process (ex. corrosion) and allows it to be controlled.

9.1. OXIDATION AND REDUCTION

The term *oxidation* has 2 meanings: **oxidation is the process through which an atom or an ion increases its positive valence or decreases its negative valence, as a result of losing a certain number (1, 2, 3) of electrons.**

The common sense (narrow) of oxidation is the chemical reactions (of combination) between oxygen and any other element and the result is an oxide. For example, when calcium and oxygen combine they form **lime**:

$$Ca + \frac{1}{2}O_2 = CaO\,.$$

In chemistry the term **oxidation** has a wider meaning: **oxidation is a chemical process through which an element losses one or more of its electrons, giving them (-i) to another element**. The element which receives electrons is not necessarily oxygen. For example the reaction between calcium and chlorine is considered to be an **oxidation reaction**:

$$Ca + Cl_2 = CaCl_2.$$

In this reaction, one calcium atom oxidizes when gives up two electrons for two chlorine atoms. Both reactions can be represented as follows:

$$Ca = Ca^{2+} + 2e^-.$$

Reduction is an opposite chemical process to **oxidation**, through which the atoms of an element receives one or more electrons from another element. For example, the reduction of copper from copper oxide:

$$CuO + H_2 = Cu + H_2O.$$

In this reaction, the positive ion of copper receives two electrons from two atoms of hydrogen. The reduction of copper in ionic form can be represented as follows:

$$Cu^{2+} + 2e^- = Cu.$$

The oxidation of an element is always accompanied by the reduction of an another element. Therefore this reaction is usually called of **oxidation-reduction** or **redox**.

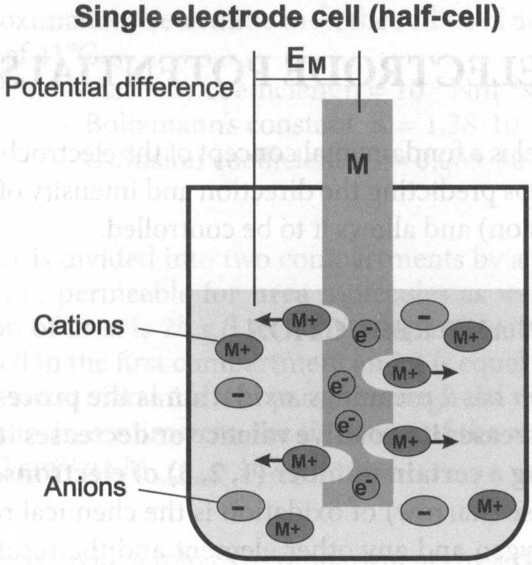


Fig. 9.1.

When a metal bar **M** (*fig. 9.1*) is placed in a solution capable to dissolve this metal, the process of oxidation begins, yielding to the metal probe electrons and forming positive ions **M**$^+$, which further will move in the electrolytic solution. As a result is generated a difference of potential E_M between the metal bar **M** and the electronic forms, which represents the potential of electrode and its value results from the **Nernst's relation**:

$$E_M = V_{metal} - V_{solution} = E_{M_0} + \frac{RT}{zF}\ln[M^{z+}].$$

Where $\mathbf{E_0}$ is a constant called **normal potential of an electrode** (is the value of E_M when $(M^{z+}) = 1$). The absolute value of potential difference can't be measured because it would require the introduction of another electrode in the electrolyte and formation of another potential difference between them. For relative measures of different metals' potential in different solution the **galvanic cell** is used.

The typical galvanic cell consists of two metal electrodes (ex. Zn and Cu) immersed in $ZnSO_4$ and respectively $CuSO_4$ solutions. For measuring the electrode potential with a voltmeter is necessary to use a reference semi cell and a measurement semi cell, connected electrically through a KCl or NH_4NO_3 bridge. Knowing the normal electrode potential allows us to predict the polarity and the electromotive force of the obtained source through galvanic cell experiment.

If two normal potential are known, respectively:

Zn/Zn^{++} –0.76 V

Cu^{++}/Cu + 0.34 V,

a galvanic cell can be used:

$$Zn/Zn^{++}/KCl/\ Cu^{++}/Cu,\quad (\textit{fig. 9.2}).$$

The solution are electrically connected through a bridge which represents a strip of filter paper moistened in KCl solution and the metallic electrodes are connected to voltmeter terminals of high resistance.

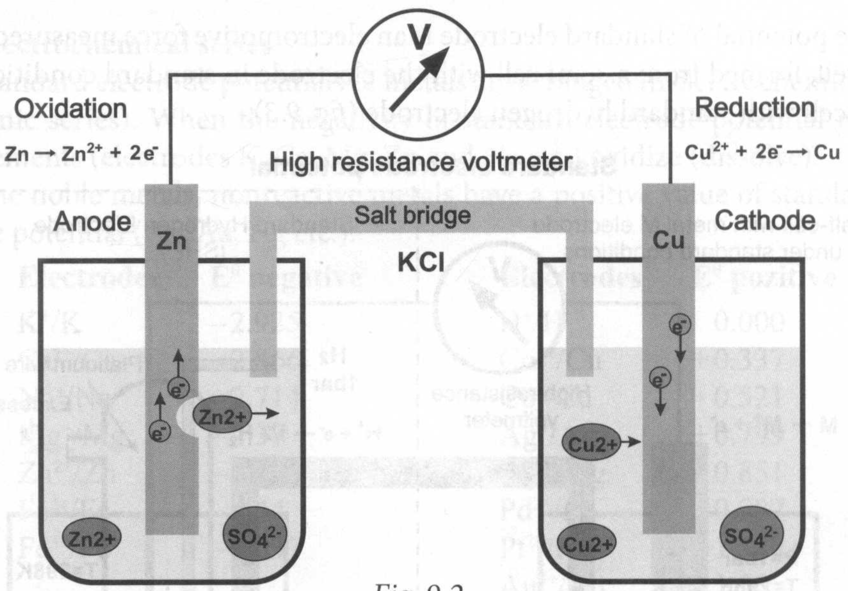


Fig. 9.2.

In this galvanic cell the zinc electrode is an anode. It oxidizes and gives up the electrons of atoms which are dissolved in an electrolyte in form of positive ions:

$$Zn = Zn^{2+} + 2e^-.$$

Copper electrode is a cathode. Dissolved copper ions receive electrons from the metallic electrode, converts to a solid metal which settles on the cathode's surface:

$$Cu^{2+} + 2e^- = Cu.$$

When the value of electric power is neglected (is used a high resistance voltmeter) is measured the potential difference between electrodes which is equal to the electromotive force of the galvanic cell.

The electromotive force is usually measured in standard conditions:

- Temperature: 25°C (298°K);
- Gas pressure: 1 bar;
- The concentration of electrolytic solution: 1 mol/dm^3.

The electromotive force measured in standard conditions denoted as E^0; it represents the sum which is resulting from the difference potential of an anode and a cathode:

$$E^0 = E_{catode} - E_{anode}.$$

For the galvanic cell previously represented:

$$E^0 = V_{Cu} - V_{Zn} = 0.34 - (-0.76) = 1.10\ V.$$

9.2. THE POTENTIAL OF STANDARD ELECTRODE

The potential of standard electrode is an electromotive force measured in galvanic cell, formed from a semi cell with the electrode in standard conditions and a semi cell with standard hydrogen electrode (*fig. 9.3*).

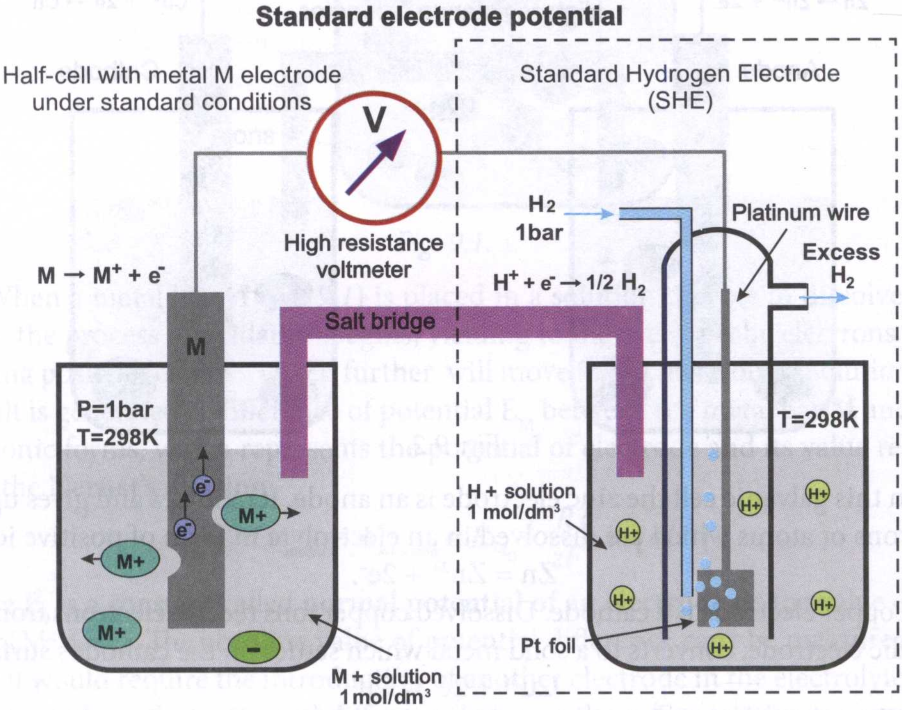


Fig. 9.3.

The standard hydrogen electrode (SHE) consists of a platinum sheet covered with black platinum (platinum dust) which is emerged in an acid solution with concentration of hydrogen ions (H^+) 1 mol/dm^3 and which is in contact with the gaseous hydrogen bubbles around electrode.

Platinum doesn't participate in the reaction but functions as a catalyst for the hydrogen redox reaction.

The potential of the standard electrode is measured relatively towards the standard hydrogen electrode, whose potential is defined as 0 volts:

$$E^0 = E^0_R - E^0(H^+/H_2) = E^0_R.$$

The metallic electrode is connected to the positive terminal of a voltmeter and the standard hydrogen electrode to the negative terminal. When a standard electrode potential is measured, the electrode is connected to the positive terminal of a voltmeter and the standard hydrogen electrode is connected to the negative terminal.

Electrochemical series

Standard electrode potentials of metals are arranged in electrochemical series (galvanic series). When the negativity of standard electrode potential is bigger, the elements (electrodes K, Ca, Mg, Zn and Na etc.) oxidize (dissolve).

The noble metals, nonreactive metals have a positive value of standard electrolyte potential (Ag, Au, Hg etc.).

Electrodes	E^0 negative	Electrodes	E^0 pozitive
K^+/K	−2.925	H^+/H_2	0.000
Ca^{2+}/Ca	−2.866	Cu^{2+}/Cu	+0.337
Na^+/Na	−2.714	Cu^+/Cu	+0.521
Mg^+/Mg	−2.37	Ag^+/Ag	+0.799
Zn^{2+}/Zn	−0.763	Hg^{2+}/Hg	+0.851
Fe^{2+}/Fe	−0.44	Pd^{2+}/Pd	+0.987
Fe^{3+}/Fe	−0.037	Pt^{2+}/Pt	+1.188
		Au^{3+}/Au	+1.50
		Au^+/Au	+1.692

9.3. CONCENTRATION CELL

This is a galvanic cell which consists of two electrodes of the same metal, placed in two solutions with different concentration of the same solvit. At temperature of 27°C for $R = 8.31\ J \cdot K^{-1} \cdot mol^{-1}$ and substituting the natural logarithm with the decimal logarithm, from Nernst's relation, we obtain:

$$E = \frac{0.06}{z} \lg \frac{[M^{z+}]_2}{[M^{z+}]_1}.$$

The positive electrode is in contact with the solution which has a higher concentration.

Applications in measuring the pH

A hydrogen galvanic cell is made. The hydrogen electrode, placed in the solution with unknown pH, through KCl bridge joins the reference hydrogen electrode placed in a solution with pH = 0. In these conditions the measured potential difference is strictly proportional to the unknown pH.

$$E = 0.06\,pH, \text{ whence } pH = \frac{E}{0.06}.$$

Practically it is more convenient to replace the reference hydrogen electrode with a reference **calomel** electrode saturated with KCl, in this case the relation is:

$$E_{relative} = E_{measured} - 0.246 = 0.06\,pH,$$

$$pH = \frac{E_{measured} - 0.246}{0.06}.$$

EXERCISES

1. The normal potential of a silver electrode placed in a silver chloride solution is +0.80 V. Determine the potential of the metallic electrode at 27°C if the solution concentration is 10^{-4} mol/dm^3.

Answer: 0.56 V.

2. The normal potential of a copper electrode placed in a $CuSO_4$ solution is +0.34 V. An electrode was places in 500 cm^3 solution prepared by mixing 250 cm^3 $CuSO_4$ 10^{-2} mol/dm^3 solutions and 250 cm^3 of water at 27°C. Determine the copper electrode potential in obtained solution.

Answer: +0.27V.

3. Through KCl bridges was made a cell between the aqueous solution A of $CuSO_4$ with 10^{-2} mol/dm^3 concentration and aqueous solution B of $CuSO_4$ with 10^{-3} mol/dm^3 concentration at temperature of 27°C. Determine the electromotive force of the cell.

Answer: 0.03V.

4. The pH of a solution of HCl with 10^{-2}M was measured by a voltmeter connected to the terminals of a circuit, which consists of a reference hydrogen electrode and an identical measured electrode. Determine the value of electromotive force E.M.F applied to the voltmeter.

Answer: 0.12V.

5. Using the results from previous exercise, determine E.M.F. measured in the circuit, where the reference electrode is a calomel electrode of saturated

KCI, and the measuring the electrode – an electrode in a solution of hydrogen 10^{-2} mol/dm^3 HCl.

Answer: 0.366 V.

6. E.M.F. measured on a hydrogen electrode relative to a calomel reference electrode with saturated KCl solution is placed in an unknown solution with $E_m = 0.546$ V. What is the pH of the solution?

Answer: 5.

7. Knowing the normal potential $Zn/Z^{++} = -0.76$ V and $Ag/Ag^{+} = +0.80$ V, determine E.M.F. a cell up of $Zn/Z^{++}|KCl|Ag/Ag^{+}$.

Answer: +1.56 V.

10. BIOELECTRICAL PHENOMENA

10.1. NERNST'S CONCENTRATION ELEMENT

The main characteristic of a living cell is the existence of an electric potential difference between the external and internal sides of the cell membrane, generated by an unequal distribution of the electric charges in those two compartments.

Within these compartments, in a stationary state, a well-determined relationship is set between all the parameters: electric potential, chemical potential, pressure etc. This fact defines **the membrane equilibrium**.

In order to reach membrane equilibrium, factors such as membrane characteristics and pecularities of the same media that it separates need to be taken into consideration.

The simplest **model of membrane biopotential source** is **Nernst's element of concentration** (*fig. 10.1*). It postulates that ionic solutions with different concentrations of the same salt are separated by a membrane with different **permeability for cations and anions**. The side with the most diluted solution have the same charge (positive or negative) as the ion with the greater permeability. In the previous figure it is presented a situation when $\mathbf{C_1 < C_2}$ and the coefficient of permeability for cations ($\mathbf{P_c}$) is greater than for anions ($\mathbf{P_a}$) is represented.

Living organisms contain a big amount of water (human organism – 60–70% from the body weight), which represents the main solvent of the polyphase disperse systems: **crystalloids** in ionized state (Na^+, K^+, Cl^-, Ca^{++}, Mg^{++}); **micromolecules; macromolecules** (in general, anions with physiological pH).

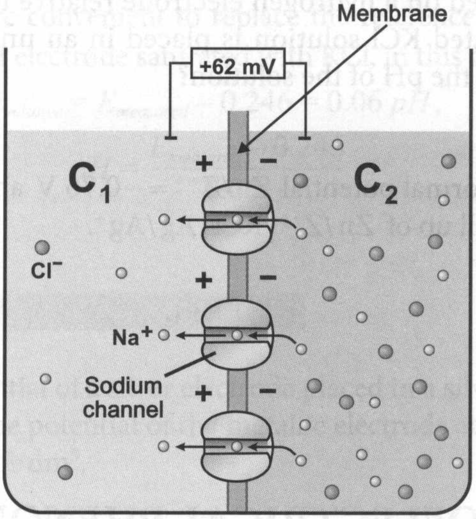


Fig. 10.1.

For each category of diffusible ions, electric potential equilibrium is defined as the potential difference between the interior (V_i) and exterior (V_e) parts of the cell membrane.

The equilibrium (V_{eq}) potential value is determined using **Nernst's equation**:

$$V_{eq} = V_i - V_e = -\frac{RT}{ZF}\ln\frac{C_i}{C_e}.$$

Where: Z – charge of the ion; F – Faraday's number; R – universal gas constant; T – absolute temperature; C_i and C_e – concentrations of ions on the respective part of the membrane.

The main ions that intervene in membrane equilibrium are: **Na^+, K^+ and Cl^-**.

In the table below are given the concentrations of these ions in the intra and extracellular mediums, as well as their equilibrium potentials, calculated according to the Nernst's relation, for the **giant axon of a squid (Loligo Paelei)**.

Ions	C_i in meq/l	C_e in meq/l	V_{eq} in mV
K^+	400	20	−75
Na^+	50	440	+55
Cl^-	52	560	−61

10.2. DONNAN EQUILIBRIUM

The majority of the cellular membranes are permeable for micro ions, such as Na^+, K^+, Cl^- etc., but impermeable for proteic (protein) macro ions. The presence of some macro ions for which the membrane is impermeable to in the system, leads to redistribution of micro ions for which the membrane is permeable, determining the appearance of an electric potential difference between the membrane's faces.

Often it is spoken about some anions that are non-diffusible, which belong to just one of the two compartments of the two compartments with selectively permeability (*fig.10.2*).

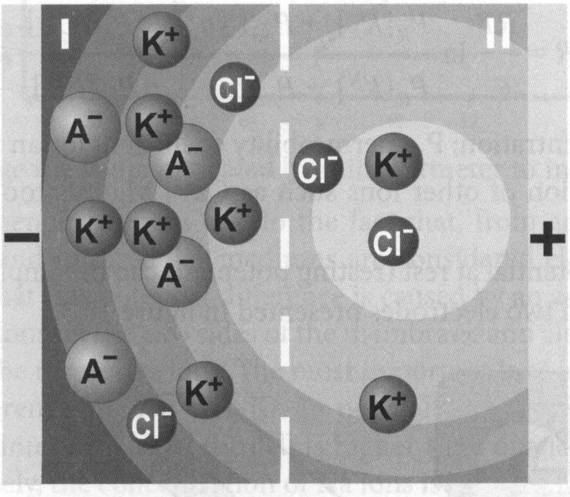


Fig. 10.2.

If non-diffusible anions with **Z** elementary charges, each with a concentration of $[A^-]$ exist only in the first compartment but concentrations of small monovalent anions and cations from those 2 compartments are $[a^-]_1$, $[C^+]_1$, $[a^-]_2$, $[C^+]_2$, than the electric neutrality of each compartment is transposed quantitatively in the following way:

$$[C^+]_1 = [a^-]_1 + Z[A^-];$$
$$[C^+]_2 = [a^-]_2.$$

In equilibrium state, in order to inhibit the flux of diffusable anions and cations, the necessary conditions lead to the following expression:

$$\Delta V = \frac{RT}{F}\ln\frac{[C^+]_1}{[C^+]_2}; \quad \Delta V = \frac{RT}{F}\ln\frac{[a^-]_2}{[a^-]_1}.$$

From these two relations results that:

$$\frac{[C^+]_1}{[C^+]_2} = \frac{[a^-]_2}{[a^-]_1} \text{ or equivalent: } [C^+]_1 \cdot [a^-]_1 = [C^+]_2 \cdot [a^-]_2.$$

This relation is called the **Donnan's equation of equilibrium.**

The respective phenomenon added to (summed with) the diffusion potentials makes a contribution (a small one) to the transmembrane potential observed in the living cell case.

10.3. RESTING MEMBRANE POTENTIAL

Theoretically, resting potential of a living cell is calculated using the **Goldman–Hodgkin–Katz equation**:

$$PR = \frac{RT}{F}\ln\frac{P_K[K^+]_e + P_{Na}[Na^+]_e + P_{Cl}[Cl^-]_i}{P_K[K^+]_i + P_{Na}[Na^+]_i + P_{Cl}[Cl^-]_e}.$$

Where: [ ] – concentration; P – permeability coefficient for an ion category.

The contribution of other ions such as Ca^{++}, Mg^{++}, is too small and can be neglected.

Membrane potential at rest (resting potential) can be simply measured with a millivoltmeter and two electrodes presented in figure 10.3.

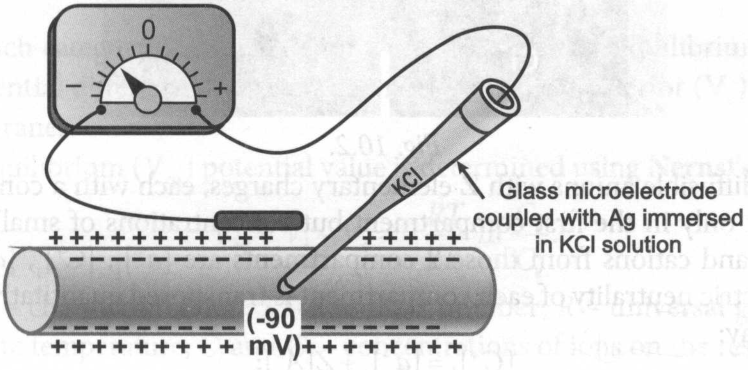


Fig. 10.3.

A glass microelectrode is a pipette obtained warming it with a tip smaller than 0.5 μm. Penetrating the membrane with the help of this microelectrode doesn't damage it significantly, excluding the occurrence of a circuit contact between the cytoplasm and extracellular fluid, causing the membrane to envelope the tip of the micropipette (because of the superficial tension). So when a microelectrode is introduced into the cell, while the other one is kept in the extracellular medium the millivoltmeter will indicate the value of membrane potential at rest (*fig. 10.4*).

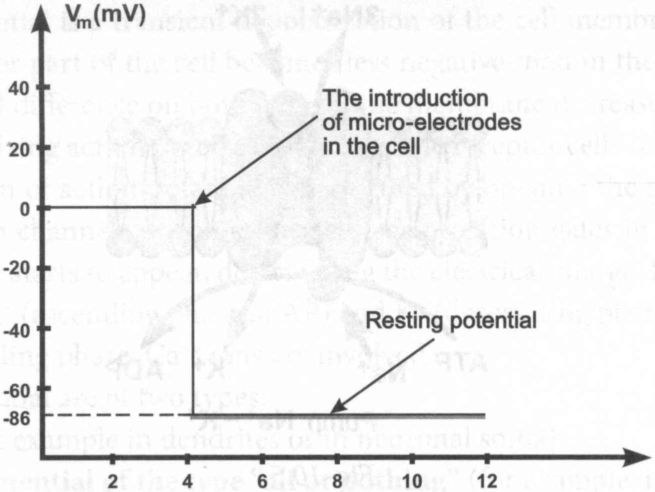


Fig. 10.4. Change the value indicated by millivoltmeter to introduction in cell.

This phenomenon happens due to the fact that, from an electrical point of view, the extra- and intracellular mediums are considered electrically neutral. It was discovered that this potential difference is caused by an asymmetric distribution of the same ions on the two sides of the membrane and also by the membrane permeability to the respective ions. The most important ions in the determination of potential difference are Na^+ and K^+. In normal conditions the concentration of K ions in the interior part of the cell is higher than outside the cell (35 times higher). Conversely, the concentration of Na ions is higher on the exterior face in comparison with the concentration on the cytoplasmic side. Despite these gradients the cell maintains its polarization by exchanging constantly ions between intracellular and extracellular environments.

There are ionic channels that allow only K^+ ion to pass in the resting state. These ensure the maintenance of membrane potential in constant limits, allowing part of K^+ ions from interior part to escape to the exterior part, but only in enough amounts to compensate the absence of Na^+. During the resting state the permeability for K^+ ions is 100 times higher than that for Na^+ ions. The membrane reaches a dynamic equilibrium by means of the mentioned ion flow, the moment when potential difference reaches a fixed value (the flow creates a potential difference of approximately 85 mV).

Also, there is a contribution from Na^+-K^+ pump (by consuming energy), which maintains the concentration gradients (*fig. 10.5*).

As a conclusion, through the cell membrane continuously take place passive ion flow aiming to reduce the electrochemical gradient but it is compensated by active flow that takes place in opposite direction, with metabolic energy consumption.

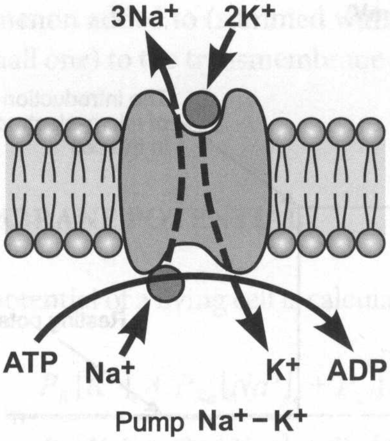


Fig. 10.5.

In order to keep their membrane potential, excitable use 60% of their energy metabolism, while other type of cells use only 30%.

For nerve cells PR = – 90 mV;
for nerve fibers – –75 mV ≥ PR ≥ – 60 mV;
for squid axon PR = – 60 mV.

The resting membrane potential is the electric potential difference between the two sides of the membrane, through which, due to all kind of ions, the total electric current annulates itself.

10.4. CELLULAR ACTION POTENTIAL

Action potentials are waves of depolarization that move along the membrane. These are an essential feature of animal life, because of their function in transmitting information in a rapid way between the tissues. Action potentials are especially used in nervous system but they can appear in the majority of cells. The investigations of these are related with giving in 1963 the Nobel Prize for Physiology and Medicine to British scientists A. Hodgkin, I. Eccles and A. Huxley for studies performed in the domain of electrophysiology of Action potentials using giant axons of Loligo Paelei.

Central and peripheral nervous systems consist of a wide network of communication in the organism, a network in which a phenomenon of bioelectric nature is used in signal transmissions, influx or nerve impulse. Nerve impulse represents the transient and propagable variation of the membrane potential of the nerve fibers called **the Action potential produced (AP)**, by a stimulus (sometimes a spontaneous cellular activity exists too).

Action potential is a transient depolarization of the cell membrane through which the interior part of the cell becomes less negative than in the resting state and the potential difference on both sides of the membrane decreases. There also exist hyperpolarizing action potentials, for example receptor cells of the retina.

The initiation of action potential is performed by opening the gates of some anion and cation channels or sometimes by closing cation gates in hyperpolarizing AP. Ion flow starts to appear, determining the electrical charge. Involved ions are especially Na^+ (ascending phase of AP) and K^+ (descending phase). In muscle cell in the ascending phase, Ca^{++} ions are involved.

Action potential are of two types:

1) **Local** (for example in dendrites or in neuronal soma);

2) **Action potential of the type "all or nothing"** (for example: in axons).

These potentials can be also measured with microelectrode experience.

Local action potential (L–AP) is obtained from depolarizing stimuli of low intensity, with a value below the so called **"threshold value"**. These **stimuli** are called **"underthreshold" stimuli**. They produce a reduced depolarization of the membrane, which propagates **decrementaly** (with losses) or **electrotonically**. For example at the giant axon a 15mV depolarization determines a local potential. The amplitude of such potential decreases, exponentially, with the distance. So local potentials are characterized by an amplitude which is proportional to the intensity of the stimulus and has decremental propagation.

Action potentials "all or nothing" (AP–an) – are initiated when the intensity of the stimulus reaches a critical value "threshold value" or "detonation threshold" (when it has liminal or supraliminal values). **"Peak potentials" (spikes)** of a nerve fiber are produced when there is an ample variation of cellular potential causing the interior part of the cell to become positive (+30mV). Its amplitude is 120mV. Once initiated, independently of the stimulus amplitude, AP–an with constant amplitude.

Therefore, AP–an is characterized by:

- constant amplitude (doesn't depend on the intensity of the (exciting element) excitant when the stimuli are luminal or supraliminal);
- propagation on large distances and without losses of speed (non-decremental);
- amplitude of peak potential, the threshold and the speed of propagation are characteristic of the fiber (or cell).

All living cells have **reactivity**, responding to the action of stimulus. The responding capacity by local action potentials, which every cell has, is called **irritability**. The responding characteristic by action potentials "all or nothing" is called **excitability** and is observed in 3 type of cells: **nerve, muscular** and **glandular**.

Active potential phases

The generation mechanism of AP–an "all or nothing" includes several complex phases presented on figure 10.6.

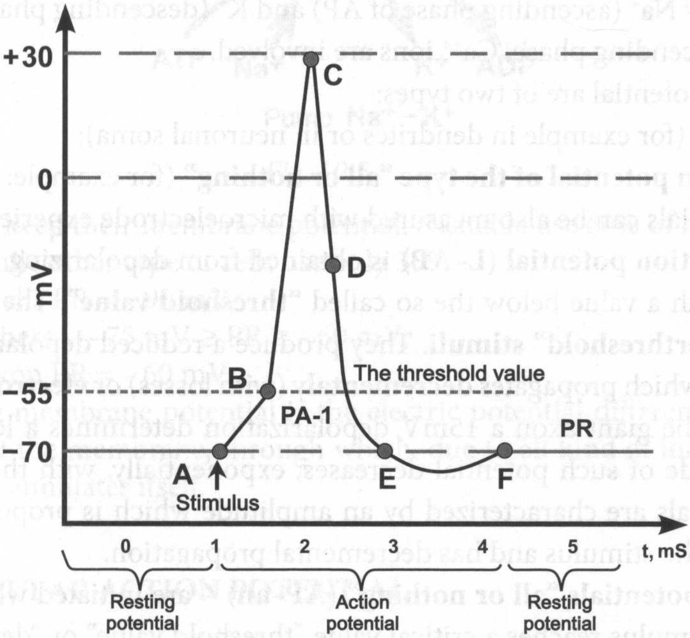


Fig. 10.6. Active potential formation in an unmyelinated nerve fiber.

AB is the **prepotential** phase – variation of resting potential, which precedes peak potential (spike) appearance.

BCD – is **the spike potential** which appears only when the stimulus rises above the threshold of excitation. It represents the fast variation of membrane potential. The amplitude of this variation doesn't depend on the intensity of stimulus. If the amplitude rises above the excitation threshold, it respects the **„all or nothing"** rule.

DE – **negative potential.**

EF – **positive potential** (traditional name).

When a long lasting stimulus is acting on a nerve fiber an accommodation occurs, which manifests by increasing the excitability threshold. The accommodation can be rapid (motor nerve fibers) or slow (sensitive nerve fibers). If the local depolarization is superior to the excitability threshold after the absolute refractory period of a spike a second one is produced and so on, the fiber presenting a repetitive response (oscillatory). Through this, a **codification in frequency** of the stimuli amplitude is performed.

AP propagation

When an action potential is produced, local modification in distribution of electron charges takes place – this modification of the polarity leads to appearance of some local electric currents between the active zone and the adjacent (neighboring) zones: **Hermann local currents**. For the appearance of a new AP, the intensity of these currents in the neighboring zones should increase above the "detonation threshold".

The propagation is performed in different ways depending on the type of fibers:

– **unmyelinated fibers**: recurrent propagation (from near to near) by local currents (Hermann), which pass the whole surface of the axonal membrane and are closed to the axoplasm and interstitial fluid.(to the center- into the exterior part, and the inverse- into the interior part) (*fig. 10.7*);

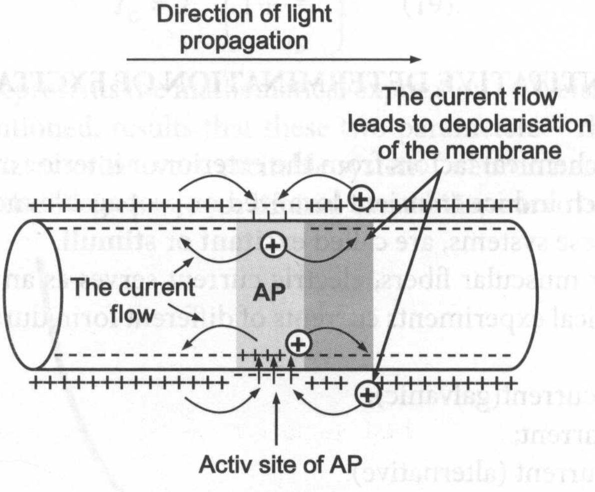


Fig. 10.7.

– **myelinated fibers**: propagation by saltatory conduction. The myelin sheath (isolator) is interrupted at the node of Ranvier, where the intra-, extracellular electric contact can be done. The local currents do not pass the whole membrane surface, but they "jump" from a node to another, close to the axoplasm and extracellular fluid (*fig. 10.8*). These currents propagate with a much higher speed and are called **Stämpfli.**

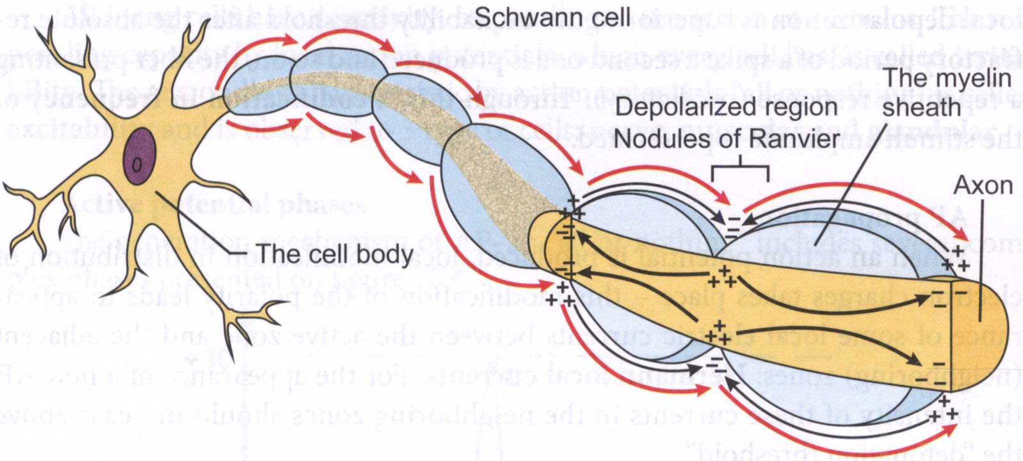


Fig. 10.8. Propagation of action potentials.

Also, the speed of propagation of AP depends on the diameter of the axon. The speed of AP in the unmyelinated axon of a squid can be 20–30 m/s when its diameter is 0.5–1mm while at a human through a myelinated axon for the same AP speed diameter is approximatively 5 micrometers.

10.5. QUANTITATIVE DETERMINATION OF EXCITABILITY

The physicochemical factors from the exterior or interior mediums of excitable systems, which induce transient localized or propagable modifications of the parameters of these systems, are called **excitant** or **stimuli**.

For nerve or muscular fibers, electric current serves as an excitatory stimulus. In physiological experiments currents of different form duration intensity are used:

– aperiodic current(galvanic);

– faradaic current;

– periodic current (alternative).

In case of rectangular electric stimuli case for activation of action potential a certain relation between liminal intensity of the stimulus (I_e) and its period (t) should exist. **Weiss,** valid with a good approximation, empirically determined this kind of relation.

$$I_e = \frac{a}{t} + b \quad (17).$$

Where: ***a*** and ***b*** – characterize the excitable system.

From this relation we can define two characteristic parameters for the excitable system: **rheobase** and **chronaxie**.

Rheobase I_r represents the minimum intensity of a stimulus with infinite duration which causes a response from the excitable system.

For an infinite duration of the stimulus ($t \rightarrow \infty$) from the relation (17) we obtain the value of the rheobase $I_r = b$ (rheobase).

Therefore, Weiss's relation, it writes:

$$I_e = I_r + \frac{a}{t_u} \qquad (18).$$

Because the useful duration, $\mathbf{t_u}$, can't be determined with precision, **Lapicque** introduced the second practical parameter, **chronaxie**. By definition, chronaxie represents the time duration, t_c, of a rectangular stimulus, with intensity equal to the double rheobase **($2I_r$)**, which induces a minimal response from the excitable system. Replacing (18) on I_e with $\mathbf{2I_r}$ and $\mathbf{t_u}$ with $\mathbf{t_c}$, we obtain: $\mathbf{a = t_c \cdot I_r}$ and it represents the liminal electric charge quantity.

Relation 18 will get its final form:

$$I_e = I_r \cdot \left(1 + \frac{t_c}{t_u}\right) \qquad (19).$$

This relation represents the mathematical expression of **Weiss–Lapicque law**.

From the mentioned, results that these two parameters – rheobase (I_r) and chronaxie (t_c) – characterize every excitable system (neuron nerve and muscle).

In figure 10.9 the Weis–Lapicque liminal excitability curve is presented.

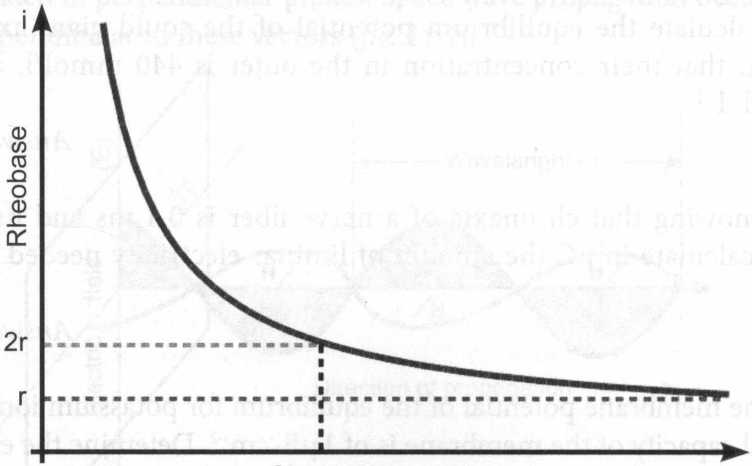


Fig. 10.9.

It is necessary to mention that for defining the time parameter (chronaxie) the double of the rheobase was used as reference but not the rheobase because as observed from the figure in this zone the slope of the curve is very small. So, it can be concluded that a small variation in intensity leads to a big duration, diminuishing the precision of measurements.

EXERCISES

1. Determine in mV, equilibrium potential for K^+ of muscle fibers at a temperature of 37°C, knowing that:
$C_e = 4 \cdot 10^{-3}\ mol \cdot l^{-1}$ and $C_i = 155 \cdot 10^{-3}\ mol \cdot l^{-1}$.

Answer: $V_{eq} = -98$ mV.

2. The equilibrium potential for the Cl^- of the squid axon membrane, at a temperature of 25°C is –58 mV. Knowing that the concentration of Cl^- in the extracellular environment is 540 $mEq \cdot g \cdot l^{-1}$. Determine the concentration of Cl^- in intracellular environment.

Answer: $C_i = 56\ mEq \cdot g \cdot l^{-1}$.

3. The resting membrane potential is –80 mV. Considering that the electric field inside the membrane is homogeneous and membrane thickness is 8 nm, determine the intensity of this field.

Answer: $E = 10^7$ V/m.

4. Calculate the equilibrium potential of the squid giant axon Na^+, if it is known that their concentration in the outer is 440 mmol/l, and inside is 49 $mmol \cdot l^{-1}$.

Answer: +55.3 mV.

5. Knowing that chronaxia of a nerve fiber is 0.4 ms and its rheobase is 35 mA, calculate in μC the amount of liminar electricity needed to induce an impulse.

Answer: q = 14 μC.

6. The membrane potential of the equilibrum for potassium ions is 134.7mV. Electrical capacity of the membrane is of $1\ \mu F \cdot cm^{-2}$. Detemine the electric charge of the membrane, with a surface area of 1 cm^2.

Answer: $1.347 \cdot 10^{-7}$ C.

7. In the thyroid gland, the intracellular concentration of I^- is 30 times higher than extracellular concentration. Express in mV equilibrium potential of iodine ions.

Answer: $V_{eq} = -60$.

8. Under the same conditions, there are two nerve fiber characterized by the same amount of liminar power (q_0), since each of the fibers has a cronaxie twice higher than that of the other, determining the ratio of their rheobases.

Answer: 1/2.

11. ELECTROMAGNETIC RADIATION. THE INTERACTION OF PHOTONS WITH MATTER

11.1. ELECTROMAGNETIC WAVES

Electromagnetic wave result from the combination of two components: an oscillating electric field, characterized by the vector $\vec{E}$, and an oscillating magnetic field characterized by the vector $\vec{H}$. The oscillations of these vectors are always in phase, situated in perpendicular planes. Space wave propagation occurs in the direction perpendicular to these vectors (*fig. 11.1*).

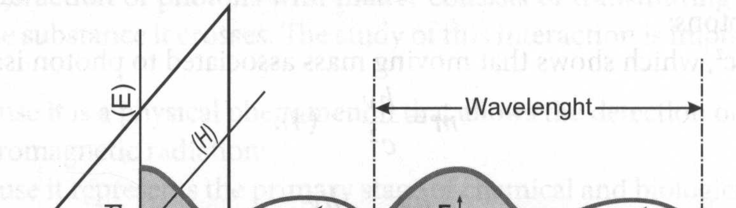


Fig. 11.1.

Electromagnetic radiation is always emitted by particles of matter, either at the level of electron or at the level of atomic nucleus and represent the mode of transporting energy in space.

Relationship of electromagnetic wave propagation is represented by this general formula:

$$A = A_0 \sin 2\pi \left(\frac{t}{T} - \frac{x}{\lambda} \right) \quad (1).$$

Where: t – time; x – distance measured in the direction of wave propagation; A – amplitude of electric field or magnetic vector at time t; A_0 – maximum amplitude; T – oscillation period; λ – wavelength.

Regardless of the origin and energy source, energy electromagnetic radiation propagates in vacuum with constant speed and has the value: **$c = 2.98 \cdot 10^8\ m \cdot s^{-1}$**.

Electromagnetic waves do not transmit energy through space continuously, but through well-defined elementary particles, called **photons (quanta)**.

Photon energy is determined from the relationship: $\varepsilon = h\nu$.

Where: $\nu = \frac{1}{T}$ – frequency in s^{-1}.

In relation to period or frequency, the wavelength is determined from:

$$\lambda = cT \text{ and } \lambda = \frac{c}{\nu} \quad (2).$$

As a particle, the photon is associated with moving mass and momentum, which result from the application of **Einstein's** equation, establishing the link between mass and energy:

$$E = mc^2 \quad (3).$$

For photons:

$h\nu = mc^2$, which shows that moving mass associated to photon is:

$$m = \frac{h\nu}{c^2} \quad (4).$$

and momentum: $p = m \cdot c = \frac{h\nu}{c^2} c = \frac{h\nu}{c}$ (5).

These characteristics can be expressed by the photon wavelength:

$$m = \frac{h}{\lambda c}; \quad p = \frac{h}{\lambda} \quad (6).$$

Our eye is able to perceive only electromagnetic radiation with wavelengths between 400 and 750 nm.

The chart below characteristics wavelenghts, and field applications of electromagnetic waves (*fig. 11.2*).

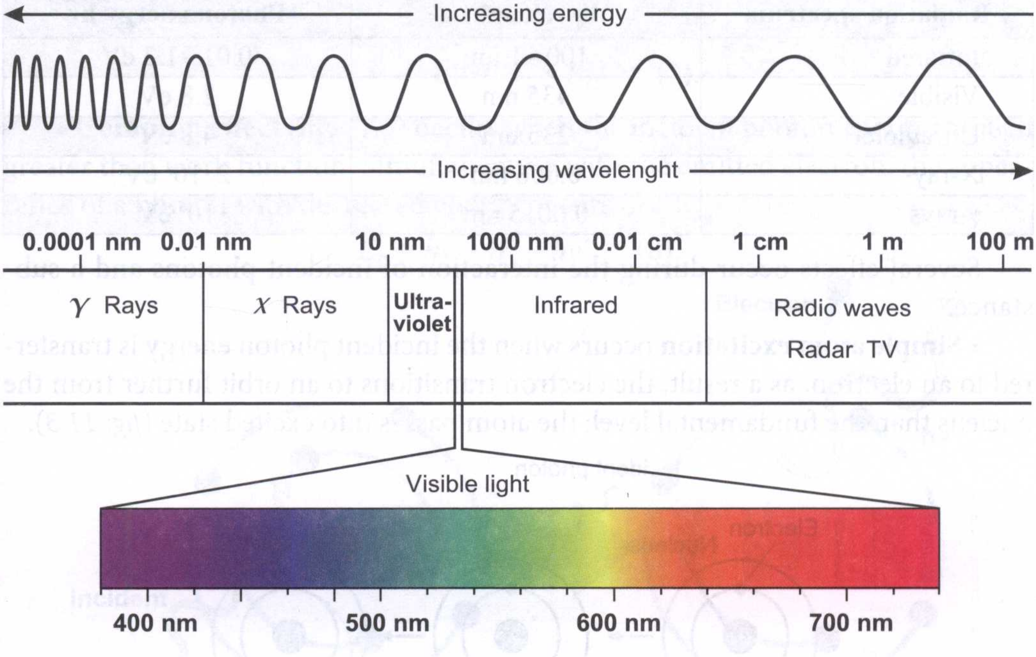


Fig. 11.2.

1.2. EFFECTS OF PHOTON AND MATTER INTERACTIONS

The interaction of photons with matter consists of transmitting their energy through the substance it crosses. The study of this interaction is important for two reasons:

- because it is a physical phenomenon that allows the detection of various electromagnetic radiation;
- because it represents the primary stage of chemical and biological actions for these radiations.

For relatively large wavelength ($\lambda > 1$ μm) photon energy does not exceed the thermal energy at room temperature. Therefore we cannot expect this radiation to produce a specific chemical or biological effect.

Specific effects of electromagnetic radiation occur only at wavelengths shorter than the lower limit of the visible spectrum, starting with ultraviolet radiation.

Concrete values of the photon energy for several electromagnetic waves, spectra are represented in the following table.

Radiation spectrum	Wavelenght λ	Photon energy hν
Infrared	100 – 1 μm	0.01 – 1.2 eV
Visible	435 nm	2.8 eV
Ultraviolet	235 nm	4.8 eV
X-rays	0.006 nm	$2 \cdot 10^5$ eV
γ-rays	0.0015 nm	10^6 eV

Several effects occur during the interaction of incident photons and a substance.

• **Simple atom excitation** occurs when the incident photon energy is transferred to an electron, as a result, the electron transitions to an orbit further from the nucleus than the fundamental level; the atom passes into excited state (*fig. 11.3*).

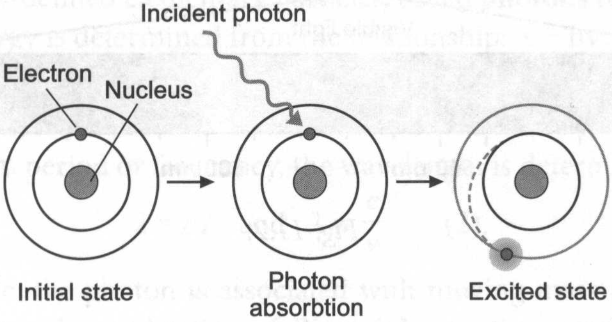


Fig. 11.3.

• **The photoelectric effect** (*fig. 11.4*) is a phenomenon of total absorption. Incident photon disappears, giving up all the energy to the electron of an atom.

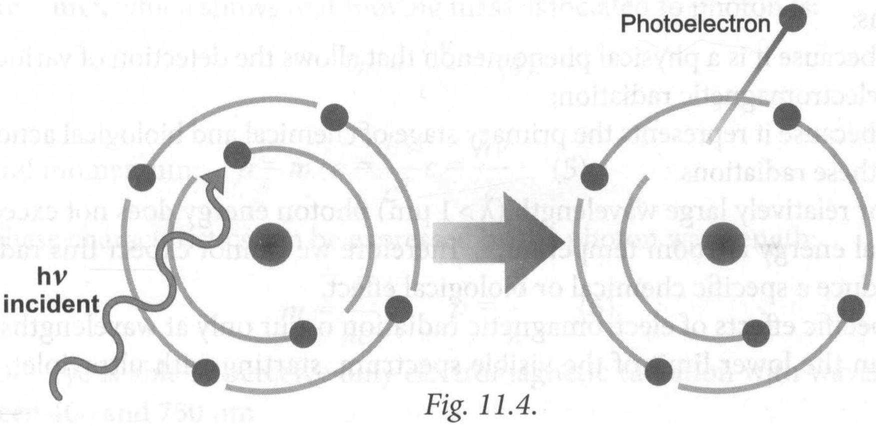


Fig. 11.4.

As a result of this energy, the electron leaves the atom with kinetic energy (**E$_c$**) equal to the difference between the incident photon energy (***hν***) and work function (**W$_i$**).

$$E_C = h\nu - W_i;$$

$$h\nu = W_i + E_C \quad (7).$$

- **Compton effect** (*fig. 11.5*) occurs when the incident photon energy is much greater than work function. Simultaneously with the emitted electron, the appearance of a photon with decreased energy occurs (*hν'*):

$$E_C = h\nu - W_i - h\nu' \quad (8).$$

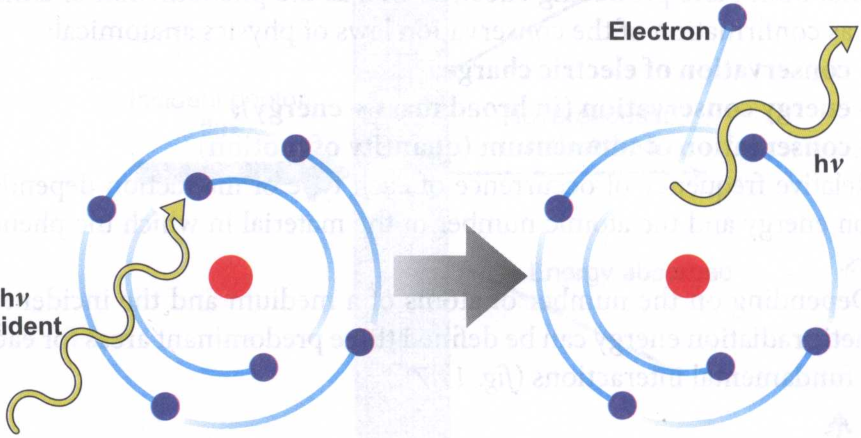


Fig. 11.5.

- **Pair creation effect** (*fig. 11.6*) – can occur when the incident photon energy is **hν > 1.022 MeV**.

In reaching the field of the atomic nucleus, the photon is explained by the nucleus, creating an ion pair consisting of a positron and electron.

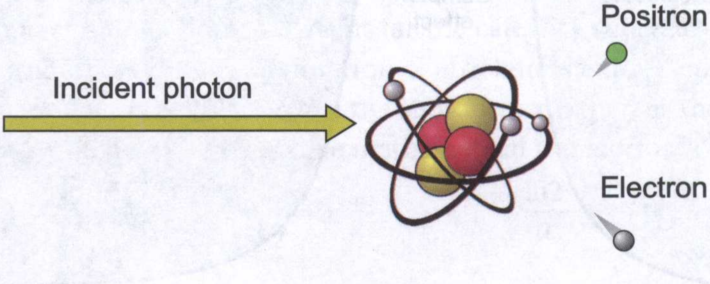


Fig. 11.6.

The formed particles possess a kinetic energy of $E_C = h\nu - 2\mathrm{m}_e$ (9).

Positron electron is the particle that differs only by the sign of electric charge, symbolically represented by **e**$^+$. In other words, the positron is the **antiparticle** **e**$^-$. When an electron and a positron collide, there is a phenomenon called **annihilation**, where both particles disappear, giving rise to two **γ** – photon energy of 0.511 MeV. Symbolic representation of this annihilation is: $e^+ + e^- \rightarrow 2h\nu$. Annihilation explains the existence of very short duration positron. Note that the term **annihilation** comes from the Latin **nihil** (nothing). The literal translation is the process in which an object was destroyed. When investigated, the particles are not actually "nothing" but transformed into another form of energy. It is important that both pairs producing effect, as well as the phenomenon of annihilation serve as confirmation of the conservation laws of physics anatomical:

– conservation of electric charge;

– energy conservation (in broad mass + energy);

– conservation of momentum (quantity of motion).

Relative frequency of occurrence of each type of interaction depends on the photon energy and the atomic number of the material in which the phenomenon occurs.

Depending on the number of atoms of a medium and the incident electromagnetic radiation energy can be defined three predominant areas for each of the three fundamental interactions (*fig. 11.7*).

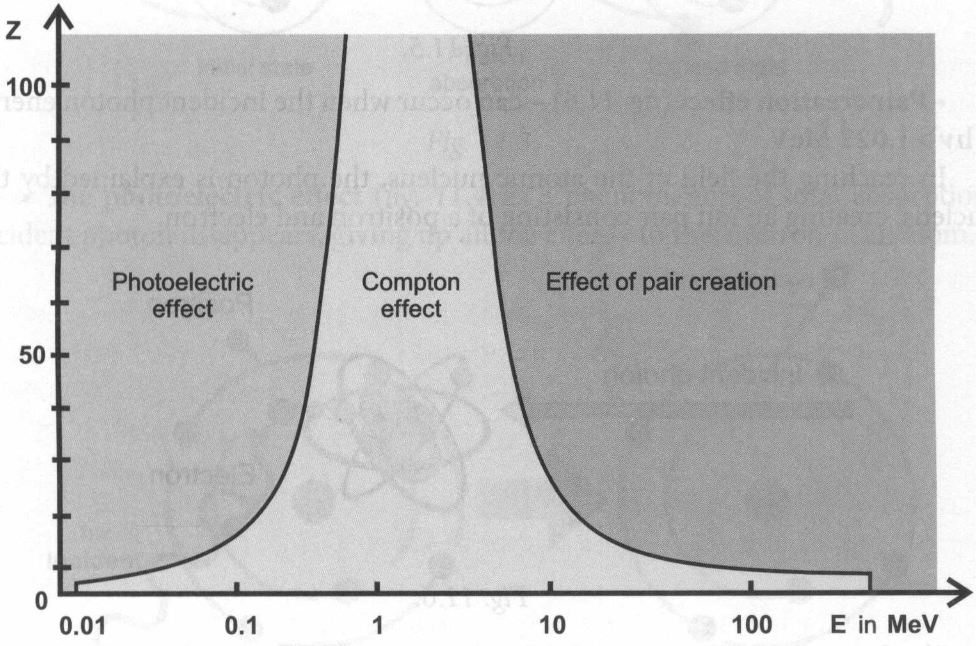


Fig. 11.7.

11.3. MITIGATION ACT

From the foregoing it follows that the energy absorption of electromagnetic radiation by the substance is a quantum phenomenon. A part of the total energy of a stream, of photons traversing a thick layer of substance **x**, the other part being absorbed or diffused (*fig. 11.8*).

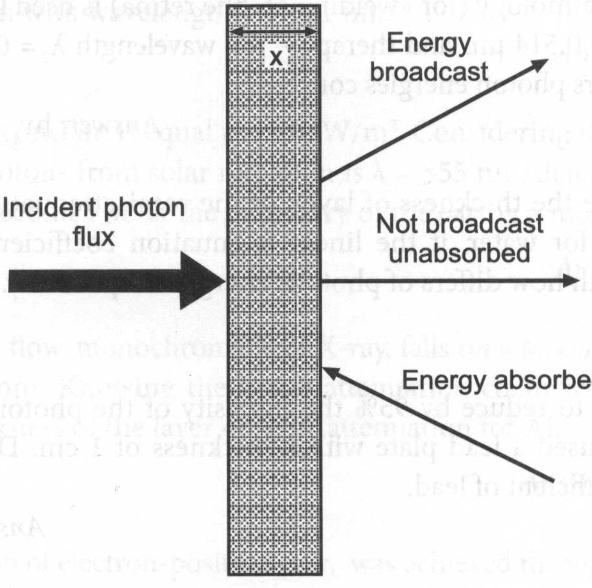


Fig. 11.8.

In the case of a monoenergetic photon flux between the number of incidents (N_0) and the number of photons that crossed layer of substance **x**, being absorbed or released (N_x), exists the relation:

$N_x = N_0 e^{-\mu x}$ (10) – mathematical expression of the law of mitigation,

where: μ – linear attenuation coefficient for the category substance on photons. If **x** is expressed in **cm**, linear attenuation coefficient is expressed in **cm^{-1}**. Between coefficient **μ** and the thickness of the substance $x_{1/2}$ there is the relationship of semi-attenuation deduced by simple mathematical transformations:

$$\frac{N_0}{2} = N_0 e^{-\mu x_{1/2}}; \quad \frac{1}{2} = e^{-\mu x_{1/2}}; \quad x_{1/2} = \frac{\ln 2}{\mu} \qquad (11).$$

EXERCISES

1. Limits the frequency of laser radiation is used in medicine, are $30 \cdot 10^9$ kHz – $15 \cdot 10^{11}$ kHz. Determine the limits of wavelength.

Answer: 1 μ – 50 μ.

2. In ophthalmology (for «welding of» the retina) is used laser radiation with wavelength $\lambda_1 = 0.514$ μm and therapy – the wavelength $\lambda_2 = 0.63$ μm. Determine how much differs photon energies concerned.

Answer: $h\nu_1 - h\nu_2 = 12.9 \cdot 10^{-20}$ J.

3. Calculate the thickness of layer, of the semi-attenuation of flow parallel of radiation γ for water, if the linear attenuation coefficient $\mu = 0.053$ cm^{-1}. Determined with how differs of photons energies respectively.

Answer: 13.1 cm.

4. In order to reduce by 95% the intensity of the photon flux of energy of 0.25 MeV was used a lead plate with a thickness of 1 cm. Determine of linear attenuation coefficient of lead.

Answer: $\mu = 3.00$ cm^{-1}.

5. It is known that a screen from lead (Z = 82) with a mass of 120 g and surface of 100 cm^2 attenuate 90% of a monochromatic X-ray flux with the wavelength of 0.69 A^0. Calculate the mass coefficient of attenuation in this screen.

Answer: $\mu/\rho = 1.92$ cm$^2 \cdot$ g^{-1}.

6. A flow of X-rays is filtered through a screen with a thickness of 0.4 cm, fabricated of an alloy composed of 70% copper and 30% aluminum. Calculate the ratio ϕ/ϕ_0 if $\mu_{Cu} = 4.5$ cm^{-1} and $\mu_{Al} = 0.5$ cm^{-1}.

Answer: $\phi/\phi_0 \approx 0.27$.

7. A protection screen was made of a lead plate with a thickness of 1 mm situated between two aluminum plates with a thickness of 2 mm each. For incident photons, semi-attenuation for lead thickness is 0.25 mm, and for aluminum – 14.8 mm. Determine the ratio ϕ/ϕ_0.

Answer: $\phi/\phi_0 = 5.19 \cdot 10^{-2}$.

8. Determine the energy of a quantum of ultraviolet radiation under the actions that skin is bronzing and is taking place if the radiation wavelength is 331 nm.

Answer: $6 \cdot 10^{-19}$ J.

9. For avulsion of electrons from the surface of metal, must to performed one work by 1.97 eV. Determine kinetic energy of the electron snatched under the action of radiation with wavelength $\lambda = 331$ nm.

Answer: $2.85 \cdot 10^{-19}$ J.

10. The solar constant is equal to 1370 W/m^2. Considering that the average of wavelength of photons from solar radiation is $\lambda = 555$ nm, determine how many photons fall per second 1 m^2 at the boundary of the earth's atmosphere.

Answer: $\approx 37.5 \cdot 10^{20}$.

11. A parallel flow, monochromatic of X-ray, falls on a screen from aluminum with thick of 5 mm. Knowing the linear attenuation coefficient $\mu = 0.35\ cm^{-1}$, calculate the thickness of the layer of semi-attenuation for Al.

Answer: $x_{1/2} = 1.98$ cm.

12. Production of electron-positron pair, was achieved through the interaction of a photon with the atomic nucleus. Knowing the photon energy is of 6.02 MeV, determine the kinetic energy of the electron.

Answer: 2.5 MeV.

13. We will consider an interaction by photoelectric effect of a photon of X-ray of 100 keV, with a zinc atom ($Z = 30$). Determine kinetic energy of the electrons emitted by on level K.

Answer: 87.76 keV.

12. X-RAY

12.1. PRODUCTION AND EMISSION. COOLIDGE TUBE

X rays were discovered experimentally in 1895 by the German physicist **Wilhelm Conrad Röentgen**. To underline the unknown nature of this highly penetrating radiation, the scientist called them **X-rays** after the discovery.

Through a number of other original experiences was found that the nature of this radiation is electromagnetic, with the wavelength of the order of **10^{-10} m.** Currently are well known not only the nature, but also their mode of production, properties, interaction with matter, application areas, but the name has remained the same – **X-rays (R·X)**.

To produce the X-rays it is necessary to bombard a metal body with accelerated electrons. The main device used for this purpose is called **the Coolidge tube**, its schematic diagram is shown in figure 12.1.

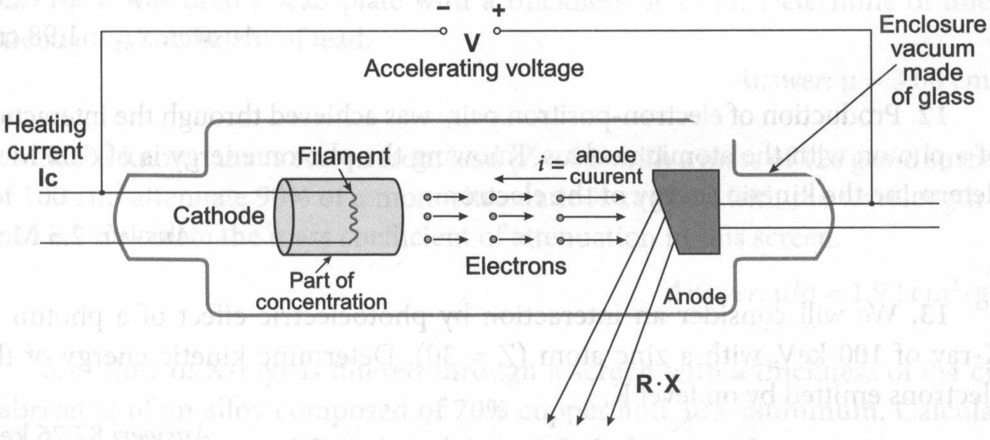


Fig. 12.1. Schematic diagram of an X-ray tube.

Electrons are emitted from the cathode filament, heated by current **Ic** with the intensity of a few amplification and accelerated toward the anode by a potential difference **V** of tens of **kV**.

The movement of electrons from the cathode to the anode corresponds to a reverse direction current with intensity of (i) a few milliamps, called **the anode current**. The intensity of this current, for the given tube, depends on the filament temperature and potential difference.

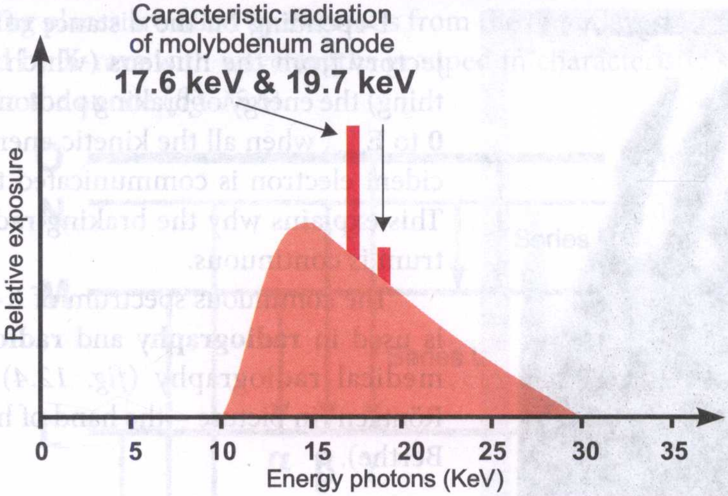


Fig. 12.2. The emission spectrum of the X-ray tube with a molybdenum anode.

If we investigate the emission spectrum of an X-ray tube (*fig. 12.2*), we find that it has the form of a continuous background, on which are overlapped several maximum levels. This double composition corresponds to a double production mechanism, **ionized** on the one hand and **braking** on the other.

Braking X-rays

When an electron with kinetic energy (**E**$_{in}$) passes near a nucleus under the force of electrostatic attraction, it changes its direction and is braked, giving rise to an X-ray photon. The energy of the photon is $h\nu = E_{in} - E_{ext}$.

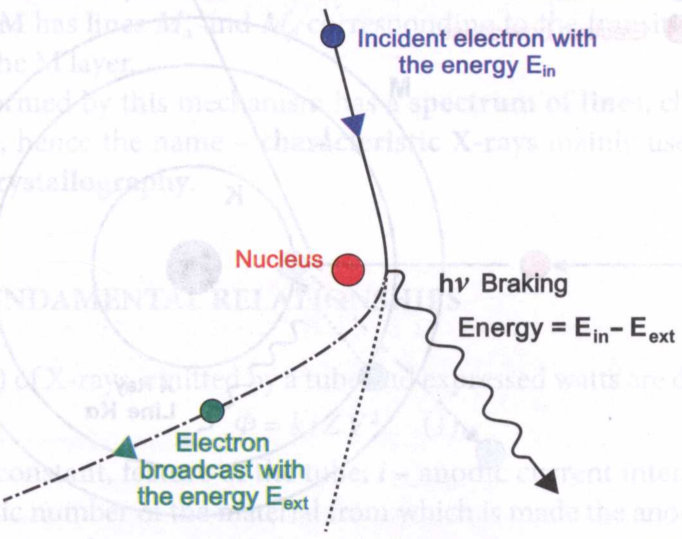


Fig. 12.3. Forming Brake X-ray.

Depending on the distance of electron trajectory from the nucleus (which can be anything) the energy of braking photons varies from **0 to E_{max}**, when all the kinetic energy of the incident electron is communicated to photon X. This explains why the braking radiation spectrum is continuous.

The continuous spectrum of X-rays braking is used in **radiography** and **radioscopy**. First medical radiography (*fig. 12.4*) belongs to Röntgen (in picture – the hand of his wife Anna Berthe).

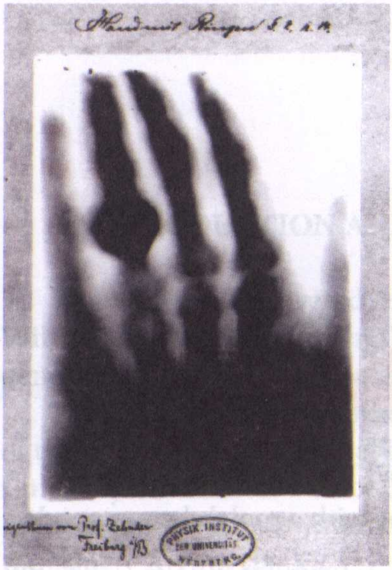


Fig. 12.4.

Characteristic X-rays

When strongly accelerated electrons (voltage V ≥ 31 800 volts) enter the anode atoms, they can snatch electrons from deep levels (K,L or M) (*fig. 12.5*).

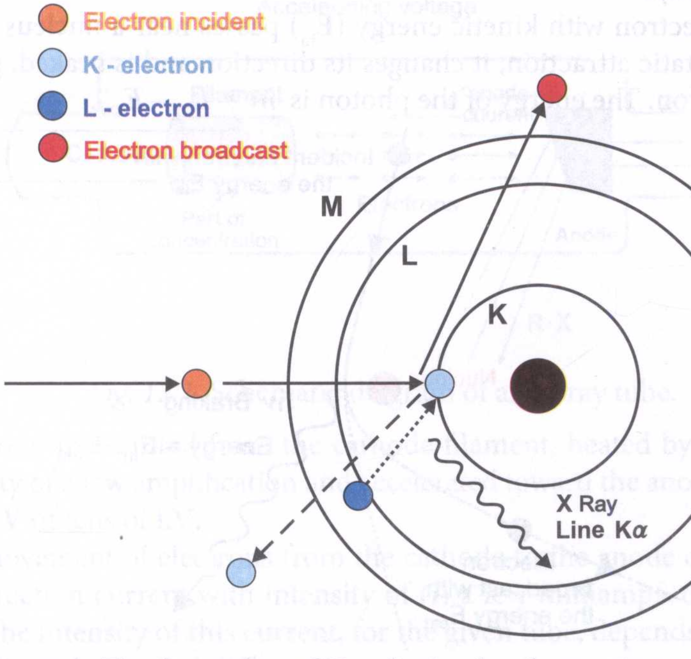


Fig. 12.5. The formation of X radiation by electronic transitions.

Remaining places are filled by electrons from the outer layers, transitions are accompanied by X-rays with wavelengths grouped in characteristic series of excited atoms in the anode (*fig. 12.6*).

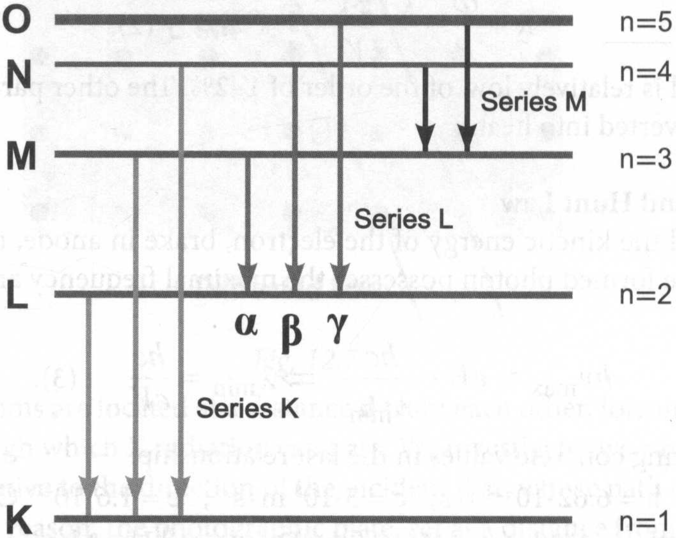


Fig. 12.6.

– Series **K** has lines K_α, K_β and K_γ corresponding to the transition from layers L, M, N to the K layer.

– Series **L** has lines L_α, L_β and L_γ corresponding to the transition from layers M, N and O to the L layer.

– Series **M** has lines M_α and M_β corresponding to the transition from layers N and O to the M layer.

X rays formed by this mechanism has a **spectrum of lines**, characteristic to anode atoms, hence the name – **characteristic X-rays** mainly used in scientific field called **crystallography**.

12.2. FUNDAMENTAL RELATIONSHIPS

Flow (**Φ**) of X-rays, emitted by a tube and expressed watts are determined by:

$$\Phi = kiZV^2 \qquad (1).$$

Where: k – constant, feature of the tube; i – anodic current intensity, in amperes; Z – atomic number of the material from which is made the anode; V – accelerating voltage, in volts.

The yield of an X-ray tube

Knowing the radiation flux (**Φ**) and current power, consumed by a functioning tube (**P = i · V**) the yield (**R**) can be determined:

$$R = \frac{\Phi}{P} = \frac{k\,i\,z\,V^2}{i\,V} = k\,Z\,V \quad (2).$$

The yield is relatively low, of the order of 1–2%. The other part of the current power is converted into heat.

Duane and Hunt Law

When all the kinetic energy of the electron, brake in anode, turns into radiation, and the formed photon possesses the maximal frequency and the minimal wavelength:

$$h\nu_{\max} = eV = \frac{hc}{\lambda_{\min}} \Rightarrow \lambda_{\min} = \frac{hc}{eV} \qquad (3).$$

Substituting concrete values in the last relationship:

$$h = 6.62\cdot10^{-34}\ \mathrm{J\cdot s};\quad c = 3\cdot10^{8}\ \mathrm{m\cdot s^{-1}};\quad e = 1.6\cdot10^{-19}\ \mathrm{C},$$

we obtain mathematical expression of the Duane and Hunt law:

$$\lambda_{\min} = \frac{1.24}{V};\ \lambda_{\min} - \text{in nm};\ V - \text{în kV} \qquad (4).$$

Mass attenuation coefficient of X-rays:

$$\mu_m = k\lambda^3 Z^3 \qquad (5).$$

Linear attenuation coefficient of X-rays:

$$\mu = \mu_m \cdot \rho \qquad (6).$$

Where: ρ – density of the absorber.

12.3. X-RAYS DIFFRACTION METHODS

In the early years of the last century, many attempts have been undertaken to determine the wavelength of the RX, but the results were unsatisfactory until German physicist **Max Laue** got the idea to use as the diffraction network crystal atoms periodic layers, the distance between which is a few angstrom. In 1912, Laue was able to get the picture of X-ray diffraction on a **calcite** crystal and determine the wavelength of the radiation.

X-ray diffraction pattern is shown in figure 12.7, points indicate ordered distribution of atoms in a simple crystal lattice.

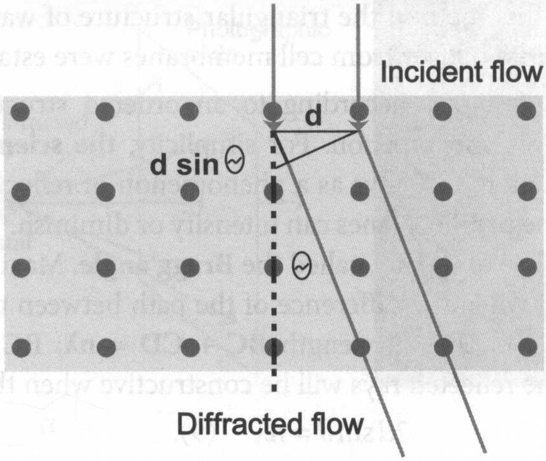


Fig. 12.7.

Layers of atoms are located at a distance **d** from each other, forming diffraction grating through which X radiation can pass. We investigate two rays diffracted at an angle **θ** relative to the direction of the incident flux, whose path difference is **d · sin θ.** For this reason, the photographic plate, set at a distance from the crystal, there will be interference picture of these two rays. Interface will be constructive when path difference contains an integer number of wavelengths of monochromatic radiation:

$$d\sin\theta = n\lambda \qquad (7).$$

and destructive, if:

$$d\sin\theta = (2\mathrm{n} - 1)\frac{\lambda}{2} \qquad (8).$$

Where: n – positive integer.

Because, unlike the two-dimensional diffraction grating, the crystal is a three-dimensional atomic system, the number of directions in space, corresponding to the constructive interface is limited. Consequently, the crystal diffraction form on the photographic plate a set of bright spots on a dark background.

This diffraction picture is called **lauegram. If the distance** (**d**) between atomic planes in the crystal is known, measuring the distances between lauegrama spots, the wavelength of the incident radiation can be determined. In practice is commonly used the reverse method: using X-ray diffraction with a knon wavelength the size and spatial network structure of the crystal can be determined.

With an accurate evaluation of a sufficient number of values **d** the strutures of various biological macromolecules such as nucleic acids, proteins, etc can be deciphered. Also by this method the triangular structure of water molecules and the structure of the lipid bilayer from cell membranes were established.

The diffraction of X-rays according to an ordered structure (crystal, biopolymer) is a diffusion phenomenon. For simplicity, **the scientist W. H. Bragg** considered the diffraction of X-rays as a phenomenon of reflection. The reflected waves from each of the parallel planes can intensify or diminish. These phenomena depend on the slip angle of the ray, called the **Bragg angle**. Maximum interference condition is achieved when the difference of the path between two rays (*fig. 12.8*) is an integer multiple of the wavelength **BC + CD = nλ; BC = CD = d·sin θ**. The interference of the reflected rays will be constructive when the path difference:

$$2d\sin\theta = n\lambda \quad (9).$$

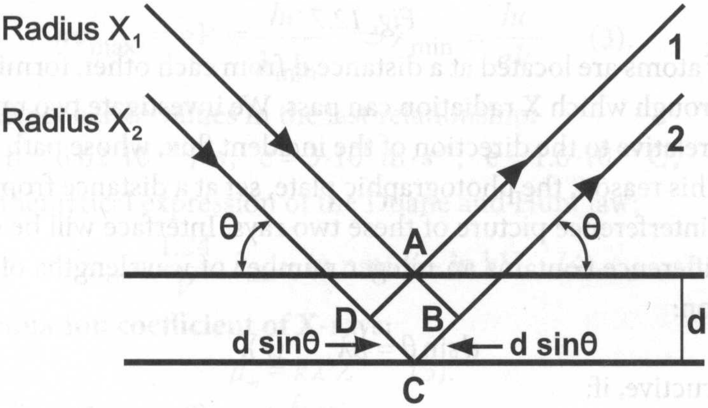


Fig. 12.8.

There are many experimental methods of X-ray diffraction, grouped in two categories of techniques:

- **Crystal cast methods;**
- **Powder methods.**

In the methods of crystal cast (diameter about 1 mm), the crystal is placed so that one of its main axis is perpendicular to the incident flow direction of the monochromatic X-rays. By rotating the crystal around this axis, rays that are falling successively on all its faces will determine the appearance of diffraction peaks (*fig. 12.9*).

Since the rotation is performed on a particular axis, the crystal surfaces have a fixed orientation and diffraction peaks appear on the photographic plate in the form of points (diffractogram or lauegram).

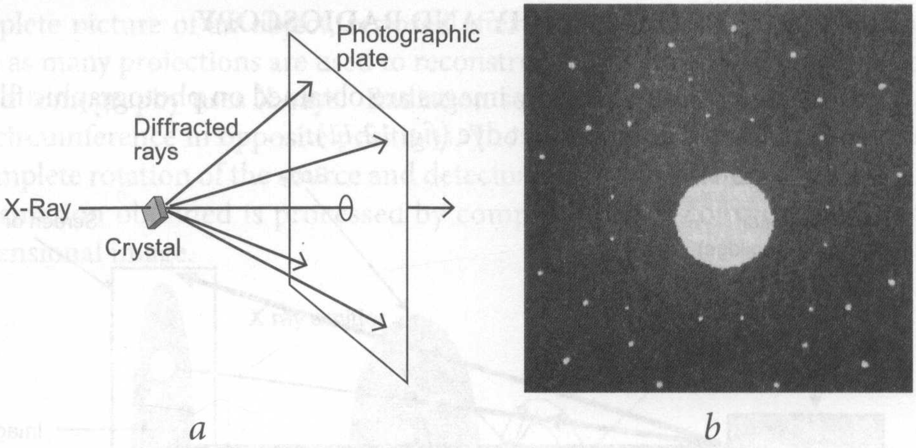


Fig. 12.9: *a* – scheme of a crystal cast method;
b – lauegram of a NaCl crystal.

When applying powder method (fig. 12.10), crystallized substance in form of powder is placed in the path of X-rays. The *entire* sample is rotated in such a way that the micro-crystals contained in powder take all possible orientations towards the incident ray, including the position when the diffraction peaks occur. Dispersed ray will form different accounts with placed axes in the direction of the incident flux. On the photographic plate, positioned perpendicular to the incident flux, the diffraction peaks will look like concentric circles.

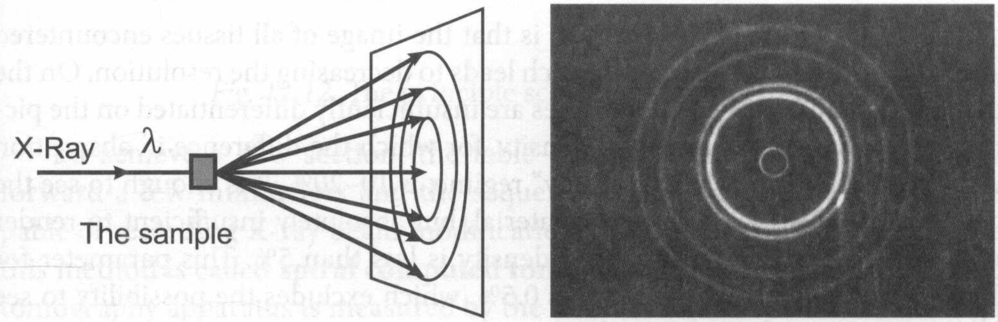


Fig. 12.10: *a* – powder method scheme;
b – *Al* polycrystal refractogram.

12.4. X-RAYS RADIOGRAPHY AND RADIOSCOPY

By radiography and radioscopy images are obtained on photographic film or on a screen containing a fluorescent dye (*fig. 12.11*).

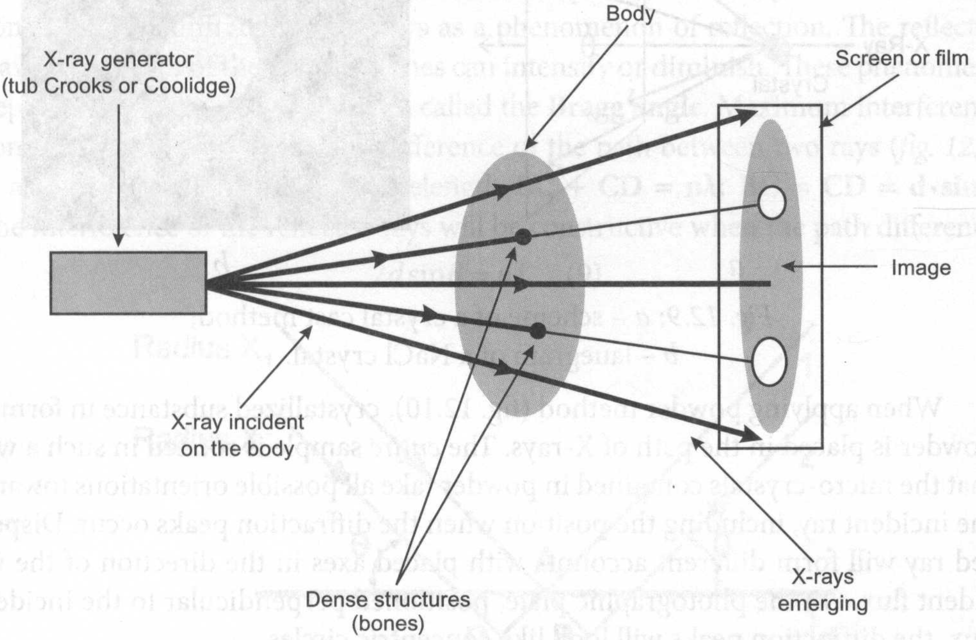


Fig. 12.11. The scheme of a X-ray radiography (radioscopy).

The *disadvantage* of this method is that the image of all tissues encountered by an X-ray overlap in the image, which leads to decreasing the resolution. On the other hand, tissues with close densities are insufficiently differentiated on the picture. The minimum difference in density, for which the difference in absorption can be reproduced in a „photography" regime, is 10–20%. It is enough to see the bones, vessels filled with contrast material, but absolutely insufficient to render soft tissues, where the difference in density is less than 5%. This parameter for white and gray matter of the brain is 0.5%, which excludes the possibility to see these structures in the radiographic image.

12.5. COMPUTERIZED TOMOGRAPHY

The term „tomography" is derived from the Greek words ***tomos*** *– section*, ***graphein*** *– write*. On the principle of operation of any computerized tomography stands the known theorem of A. Einstein, according to which the true and complete picture of an object, restored after his projections, is more extensive, with as many projections are used to reconstruct it. As general radiography, computed tomography uses X-rays. Radiation source and collector are arranged on the circumference in opposite positions. The image of a section is obtained after a complete rotation of the source and detector around the patient (*fig. 12.12*). The information obtained is processed by computer and is converted into a two-dimensional image.

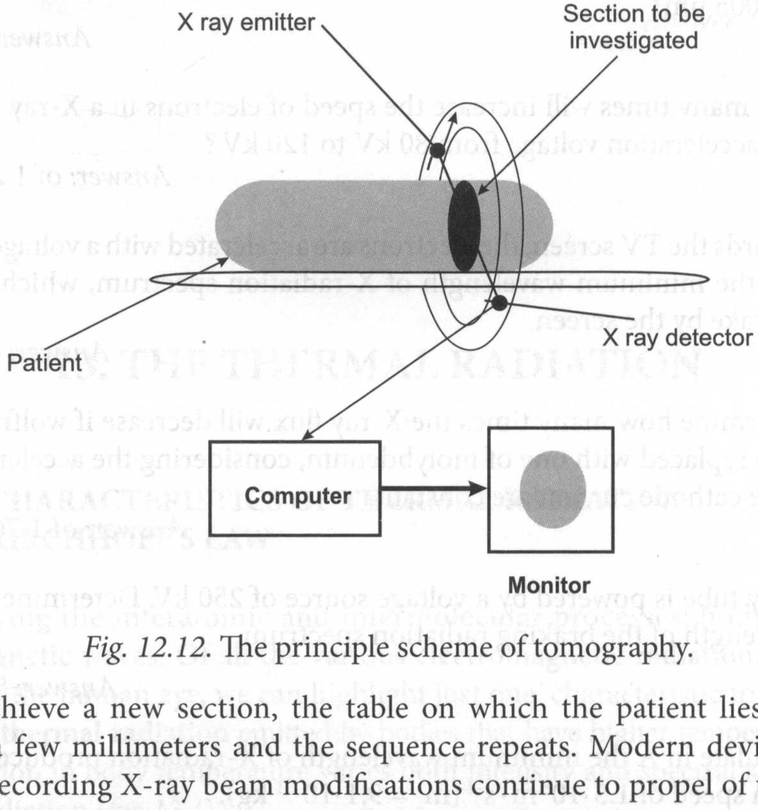


Fig. 12.12. The principle scheme of tomography.

To achieve a new section, the table on which the patient lies is pushed forward a few millimeters and the sequence repeats. Modern devices are capable of recording X-ray beam modifications continue to propel of the patient, this method is called **spiral computed tomography.** Performance of computed tomography apparatus is measured by the number of used detectors: the higher the number is, the more efficient is the appliance. To enhace the contrast of less dense tissues ion-containing contrast agents are used (radiation absorber X), administered intravenously or orally.

EXERCISES

1. How many times X-ray light wave, which has a photon energy of 59 keV is lower than violet radiation with a wavelength of 400 nm?

Answer: of 16 000 of times.

2. What is the energy of a photon of monochromatic radiation with a wavelength of 0.005 nm?

Answer: $4 \cdot 10^{-14}$ J.

3. How many times will increase the speed of electrons in a X-ray tube at an increase of acceleration voltage from 80 kV to 120 kV?

Answer: of 1.2 of times.

4. Towards the TV screen, the electrons are accelerated with a voltage of 20 kV. Determine the minimum wavelength of X-radiation spectrum, which occurs at electrons brake by the screen.

Answer: 0.062 nm.

5. Determine how many times the X-ray flux will decrease if wolfram anode tube will be replaced with one of molybdenum, considering the accelerating voltage and the cathode current are constant.

Answer: of 1.76 of times.

6. X-ray tube is powered by a voltage source of 250 kV. Determine the minimum wavelength of the braking radiation spectrum.

Answer: $5 \cdot 10^{-3}$ nm.

7. Calculate in Å the minimum wavelength of X-radiation produced by electrons with a speed of $1.5 \cdot 10^{8}\ \mathrm{m \cdot s^{-1}}$ ($m_e = 9.1 \cdot 10^{-31}$ kg).

Answer: $\lambda = 0.19$ Å.

8. The energy level (K) of the anode atoms is –69 keV. Calculate the energy range (L_α), corresponding to the electron transition from the M on the L level.

Answer: 9.58 keV.

9. The yield of X-ray tube is 2%. Calculate the X-ray flow (Φ) provided if the tube operates at a voltage of 50 kV and at anode current of 3 mA.

Answer: $\Phi = 3$W.

10. When passing through a bone tissue with a thickness of 20 mm, the X-ray flux was attenuated twice. Determine the linear attenuation coefficient.

Answer: $\mu \approx 0.035\ \text{mm}^{-1}$.

11. An X-ray point source emits N photons per second. When removing the source the flow describes a cone. Knowing the uniform flow at a distance of 1 cm from the source (Φ_0), determine uniform flow (Φ) the distance *d* cm from the source.

Answer: $\Phi = \Phi_0/d^2$.

13. THE THERMAL RADIATION

13.1. CHARACTERISTICS OF THERMAL RADIATION. KIRCHHOFF'S LAW

Following the interatomic and intermolecular processes, bodies radiate electromagnetic waves. Of all the various electromagnetic radiation, visible or invisible to the human eye, we can highlight just one, characteristic to all bodies. This is the **thermal radiation** emitted by bodies that have higher temperature than 0°K. Variation in body temperature varies both intensity and spectral structure of thermal radiation (*fig. 13.1*).

We call **radiation flux (Φ)** the average power of the radiation. In the international system it is measured in watts (W). The flux of radiation emitted by 1 m^2 of body surface is called **energetic radiant (R_e)**, expressed in W/m^2.

The ability of a body to absorb the energy of electromagnetic radiation is characterized by the **absorption coefficient (α)** equal to the ratio of the absorbed and incident radiation flux:

$$\alpha = \frac{\Phi_{abs}}{\Phi_{inc}} \leq 1 \quad (1).$$

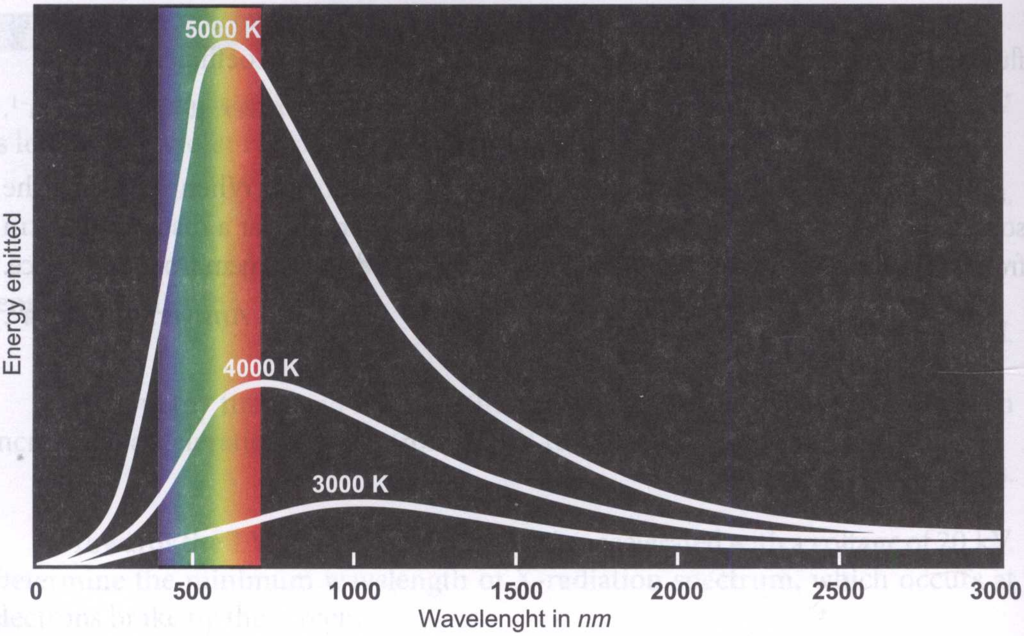


Fig. 13.1.

Energetic radiant of a body in a small range of wavelengths ($\Delta\lambda$) is expressed by the equation:

$$\Delta R_e = r_\lambda \Delta\lambda \quad (2).$$

Where: $\mathbf{r}_\lambda$ is called the **spectral density of energy radiation** of the body.

Bodies whose absorption coefficients are equal to one unit, for radiation of any wavelength, are called **black bodies** (*fig. 13.2*). Body whose absorption coefficients are less than one unit, but does not depend on the wavelength of the incident radiation, are called **gray bodies.**

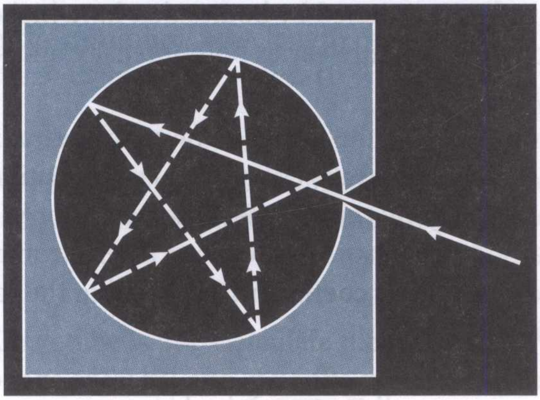


Fig. 13.2. The model of a black body.

In nature does not exist absolute black or gray bodies. These notions are physical abstractions, necessary to express mathematically certain regularities that, with some approximation, in specific cases can be used for real bodies. For example, in some researches gray bodies are considered those bodies which have an absorption coefficient of about 0.9 of infrared light.

Quantitative correlation between the emission and absorption of electromagnetic energy was established by **Gustav Kirchhoff** in 1859.

At constant temperature, the ratio of the spectral density of the energetic radiant and monochromatic absorption coefficient is the same for all bodies, including absolutely black bodies:

$$\left(\frac{r_\lambda}{\alpha_\lambda}\right)_1 = \left(\frac{r_\lambda}{\alpha_\lambda}\right)_2 = \ldots = \frac{\varepsilon_\lambda}{1} \quad (3).$$

Where: ε_λ is the spectral density of energetic radiant of the black body. Mathematically, this law can be expressed by the relation:

$$\frac{r_\lambda}{\alpha_\lambda} = \varepsilon_\lambda \quad \text{sau} \quad r_\lambda = \alpha_\lambda \cdot \varepsilon_\lambda \quad (4).$$

13.2. THERMAL RADIATION LAWS FOR BLACK BODIES

Stefan-Boltzmann law reads as follows: **energetic radiant of a black body is directly proportional to the fourth power of the absolute temperature of the body:**

$$R_e = \sigma T^4 \quad (5).$$

Where: σ is the Stefan–Boltzmann constant:

$$\sigma = 5.67 \cdot 10^{-8} \; W/m^2 \cdot K^4.$$

For gray bodies:

$$R_e = \alpha \sigma T^4 \quad (6).$$

Wien's law: the wavelength λ_{max}, which has the maximum value of spectral density of the energetic radiant of a black body, varies inversely proportional to the absolute temperature of that body.

$$\lambda_{\max} = \frac{b}{T} \quad (7).$$

Where: $b \approx 0.2898 \cdot 10^{-2} \; m \cdot K$ is Wien's constant.

Wien's law is also called the **displacement law**, because according to it at the temperature increase of the black body the maximum spectral density of the energetic radiant is shifted to shorter wavelengths.

Using quantitative spectral analysis, on the basis of this law we can determine the temperature of distant bodies, the method is called **optical pyrometer**.

On the other hand, if the body temperature is known, we can determine the wavelength with the maximum spectral density of energetic radiant. Based on this law λ_{max} wavelength of electromagnetic radiation emitted by the human body can be determined, the skin temperature is considered to be of 32°C.

$$\lambda_{\max} \approx \frac{29 \cdot 10^{-6}\, m}{3.05} = 9.5\, \mu m$$ – radiation in the infrared.

We mention that to maintain constant temperature, the human body gives away thermal energy into environment through four ways:

- Thermal conductivity;
- Convection;
- Evaporation;
- Electromagnetic radiation.

It was found that, in a relaxed state, at the ambient temperature of 20°C, through radiation human body gives 66% of the total energy and the same conditions, performing strenuous exercise, only 15%.

13.3. SUN RADIATION

The most powerful source of thermal radiation, which conditions life on Earth, is the Sun.

Some characteristics of the sun:

- appearance – about 4.59 billion years ago;
- diameter ≈ $1.4 \cdot 10^6$ km;
- mass of the sun ≈ $2 \cdot 10^{30}$ kg;
- average density of $1.408 \cdot 10^3$ kg/m^3;
- gravitational acceleration 274 m/s^2;
- chemical composition of the photosphere (in %):

hydrogen – 73.46	iron – 0.16
helium – 24.85	sulphur – 0.12
oxygen – 0.77	neon – 0.12
carbon – 0.29	nitrogen and others – 0.21;

- surface temperature – 5800°K;
- core temperature ≈ $15 \cdot 10^6$°K;
- total radiation – $3.846 \cdot 10^{26}$ W.

Solar power is the thermonuclear conversion reaction of hydrogen in helium:

$$ {}_{1}^{2}H + {}_{1}^{3}H \rightarrow {}_{2}^{4}He + {}_{0}^{1}n . $$

Earth, which is at a distance of $1.496 \cdot 10^{11}$ m, receives $2 \cdot 10^{17}$ W of solar energy.

Physical systems realize an energy exchange with the environment and with other systems. A system that receives more energy than gives – warms. A system that gives more energy than receives – cools. Earth receives energy in form of radiation and also gives energy in the universe (*fig. 13.3*).

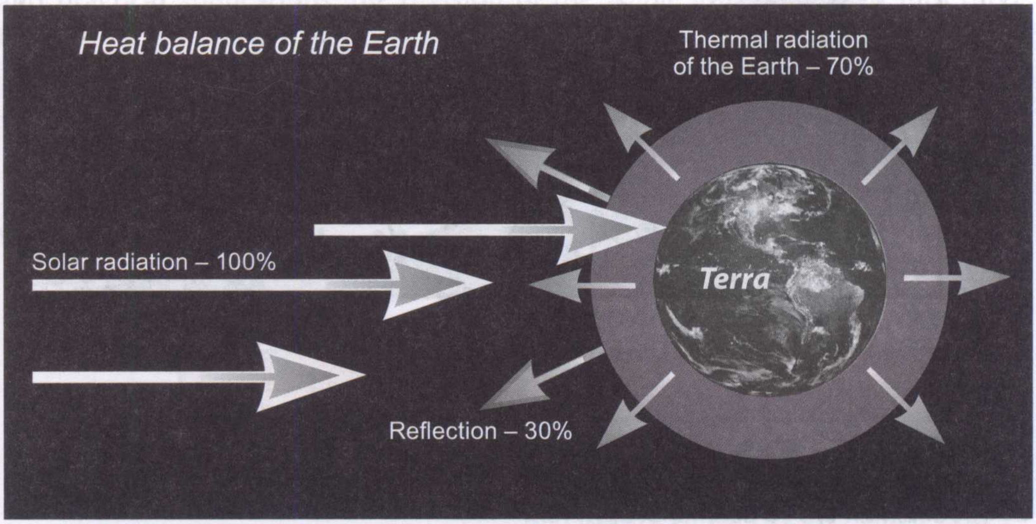


Fig. 13.3.

The energy balance of the earth is given by the equality between the energy received from the sun and the energy given into space.

Earth absorbs 70% of the energy received from the sun, and reflects 30% back into space.

Being a warm body, the earth radiates into space thermal energy, equivalent to those absorbed 70%. The thermal balance keeps constant earth's average temperature over time.

Flux of solar radiation, per 1 m^2 of atmosphere border area is **1 370 W**, called **the solar constant** and only **1 000 W/m^2** reach earth's surface.

Attenuation of solar radiation in the atmosphere causes changes in its spectral structure. In figure 13.4 the spectrum of solar radiation at the atmosphere border and on earth's surface when the sun is in zenith is shown.

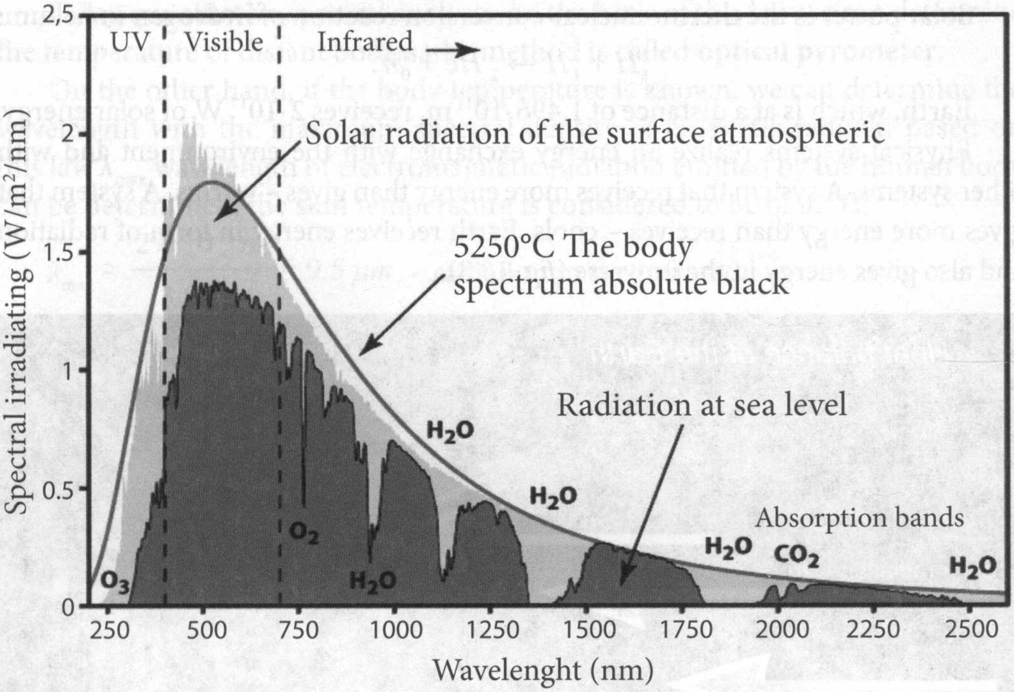


Fig. 13.4.

Atmospheric surface solar radiation spectrum corresponds largely to black body radiation spectrum, with $\lambda_{max} = 470$ nm.

At the sea level the spectrum has $\lambda_{max} \approx 555$ nm and contains more absorption bands, depending on the wavelength. The main ultraviolet absorbing substance is ozone, for infrared radiation it is water; for microwaves are water and carbon dioxide.

Dosed solar radiation is applied in medicine as a therapeutic factor (**heliotherapy**) and as a hardening factor of the body.

13.4. INFRARED RADIATION (IR). ITS PROPERTIES AND EFFECTS ON LIVING ORGANISMS

In 1800, English astronomer **Sir William Herschel** studied the effects of heat produced by the sun's rays. For this he used a glass prism. By moving a thermometer along the formed spectrum, he managed to measure the heating effect of different colors, from violet region until the red end of the spectrum. Moving the thermometer beyond the red end of the spectrum, where there is no visible light, the heating effect continued to grow. Thus, he discovered a new region of the electromagnetic spectrum – **infrared radiațion**.

IR radiations, like all electromagnetic radiations, are able to reflect, refract, undergoes diffraction, interference and polarization. In relation to visible and ultraviolet radiation, the IR has **a higher propagation velocity in different environments and a more pronounced thermal effect**.

IR fields, conventional, starts at 760 nm and extends to the limit of wavelengths up to 343 μm, where microwaves start (*fig. 13.5*). Depending on the wavelength and practical applications, the division of the whole area is allowed into three types: IR-A; IR-B and IR-C.

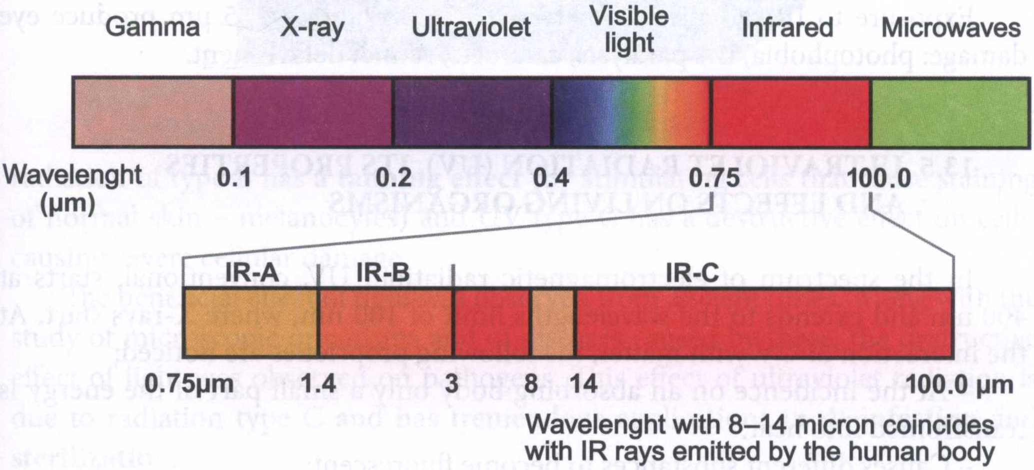


Fig. 13.5.

From the broad spectrum of infrared radiation groups of from 0.75 to 400 μm were studied, which, according to medical use, can be classified as follows:

– IR from the visible spectrum limit up to 6 μm are used for therapeutic purposes, from which only those with $\lambda < 1.5$ μm are penetrating radiation (artificial or solar);

– IR with $\lambda > 6$ μm include radiation emitted by human body, ground and everything around us.

The effect of Son's IR radiation on the human body is manifested indirectly by the modification of the thermal gradient of the skin.

In therapy, depending on the skin permeability is used the following classification of IR radiations:

– with $\lambda > 1.5$ μm are absorbed on the surface;

– with $\lambda > 1.5$ and $\lambda < 5$ μm are absorbed by epidermis and dermis;

– with $\lambda > 0.75$ μm and $\lambda < 1.5$ μm are penetrated depending on pigmentation and temperature etc.

IR radiation stimulates skin activity, making it more sensitive to external excitation, but can calm pain, either by direct inhibitory action on the affected nerves, or by action on the circulatory system.

IR radiations affect the circulation as follows:

– Through vasodilatation, intensifying exchanges between cells by enhancing osmotic phenomena and increasing blood flow;

– Through the blood and nervous systems reactions, acting on glandular secretions and general metabolism.

On the other hand, through the strong irradiation of cells with IR radiation $\lambda < 1.5$ destruction occurs.

Exposure to IR radiation with IR $\lambda > 0.75$ μm and $\lambda < 1.5$ μm produce eye damage: photophobia, iris paralysis, cataracts, retinal detachment.

13.5. ULTRAVIOLET RADIATION (UV). ITS PROPERTIES AND EFFECTS ON LIVING ORGANISMS

In the spectrum of electromagnetic radiation, UV, conventional, starts at 400 nm and extends to the wavelengths limit of 100 nm, where X-rays start. At the interaction of UV with matter, the following proprieties are noticed:

– At the incidence on an absorbing body only a small part of the energy is transfromed into heat;

– Causes different substances to become fluorescent;

– Have a photoelectric effect;

– May cause oxidation, reduction, polymerization, photolysis and biochemical reactions.

UV radiation has special biological effects, justifying their use for therapeutic purposes by direct irradiation or after previous administration of an irradiated surface with photosensitizing substances.

UV irradiation of living organisms changes in calcium and phosphorus levels of blood. An increase concentration of calcium and phosphorus in plasma and a decrease concentration of blood sugar is observed. These radiations activate circulation and increase red blood cell ability to fix oxygen. Lowers blood pressure especially at hypertensive patients.

There are three types of ultraviolet radiation that differ in wavelength and have different effects on living organisms: UV type A – from 400 nm to 315 nm; UV type B – from 315 nm to 280 nm and UV type C – the 280 nm to 100 nm (*fig. 13.6*).

These three types of ultraviolet radiations have different effects on organisms. Type A UV radiation has an erythematous effects (of inflammatory type), UV

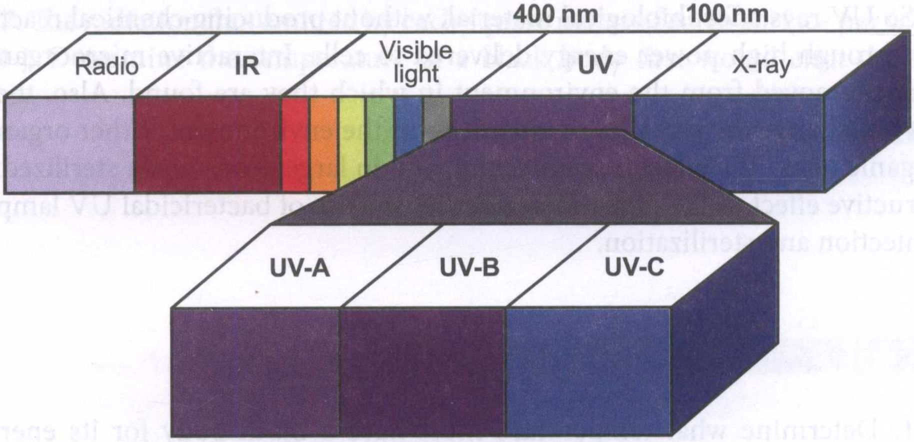


Fig. 13.6.

radiation of type B has a **tanning effect** (by stimulating cells that cause staining of normal skin – melanocytes) and UV type C has a destructive effect on cells, causing severe cellular damage.

The beneficial effect of light was observed from ancient times. Along with the study of microscopic organisms and of diseases caused by these, the destructive effect of light was observed on pathogens. This effect of ultraviolet radiation is due to radiation type C and has tremendous applications in **disinfection** and **sterilization**.

To kill microorganisms, UV rays of type C penetrate cell membrane, traverse the cell contents and destroy cellular DNA (*fig. 13.7*), causing damage that prevents bacterial activity and its ability to reproduce.

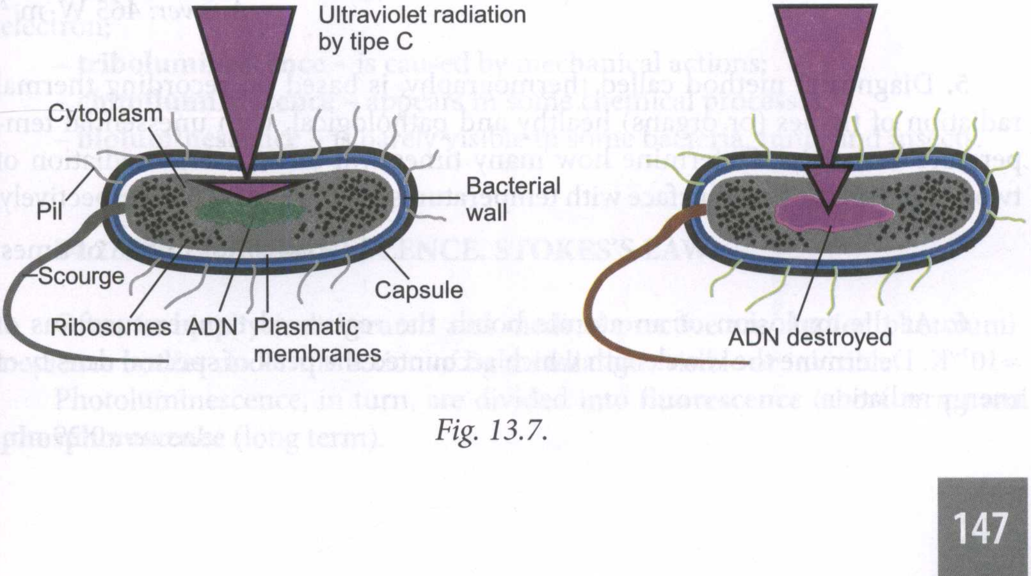


Fig. 13.7.

So UV rays affect biological material, without producing chemical reactions, only through high power energy delivered to cells. Interactive microorganisms are not removed from the environment in which they are found. Also, the UV does not change the particles or chemicals in the environment, either organic or inorganic ones. The effect is disinfectant, and in large dose – even sterilized. The destructive effect of UV radiation underlies the use of bactericidal UV lamps for disinfection and sterilization.

EXERCISES

1. Determine what temperature must have a black body for its energetic radiant to become equal to 500 W/m^2.

Answer: 306°K.

2. The absorption coefficient (α) of a gray body is 0.5. At what temperature the energetic radiant of this body will be equal to 500 W/m^2?

Answer: 364°K.

3. Surface of a black body is $T = 1000$°K. How many times will increase the radiation power of the body, if one half of its surface will be heated and the other half will be cooled at $\Delta T = 100$°K?

Answer: of 1.06 of times.

4. Determine the energetic radiant of the human body at 36°C, considering it a gray bodies with the absorption coefficient (α) equal to 0.9.

Answer: 465 $W \cdot m^{-2}$.

5. Diagnostic method called thermography, is based on recording thermal *radiation of tissues (or organs) healthy and pathological, with unessential* temperature variations. Determine how many times will differ energy radiation of two sectors on the body surface with temperatures of 30.5 and 30.0°C respectively.

Answer: of 1.0068 of times.

6. At the explosion of an atomic bomb the registered temperature was of $\approx 10^7$°K. Determine the wavelength which accounted the peak of spectral density of energy radiation.

Answer: 0.29 nm.

7. The radiation flux emitted by each square meter of earth surface per second is 90 J. Determine the temperature of a black body that would emit the same energy.

Answer: 200°K.

14. LUMINISCENCE

14.1. TYPES OF LUMINESCENCE

Luminescence represents the additional radiation of the body in relation with its thermic radiation at a given temperature witch takes a much greater period than the period of the light waves.

Thermic radiation in a visible domain occurs when the temperature of a radiant body is hundreds or thousands of degrees, while he can be luminescent at any temperature. For this reason, luminescence often is called **cold radiation.**

For the occurrence of luminescence a source of energy is needed, that is different from thermal energy balance that corresponds to the given temperature. Types of luminescence:

– **photoluminescence** – appears from absorption of the light;

– **electroluminescence** – is conditioned by the electric field in which the substance is placed;

– **chatodoluminescence** – is caused by irradiation of substance with fast electron;

– **triboluminescence** – is caused by mechanical actions;

– **chimiluminescence** – appears in some chemical processes;

– **bioluminescence** – is barely visible in some bacteria, fungi and insects.

14.2. PHOTOLUMINESCENCE. STOKES'S LAW

Often in biophysical research and medical practice are studied photoluminescence bodies under the action of light radiation, visible or invisible.

Photoluminescence, in turn, are divided into **fluorescence** (short-term) and **phosphorescence** (long term).

After the mechanism of elementary processes photoluminescence is certified in: resonance, spontaneous, stimulated and of recombination.

The initial act of photoluminescence is the excitation of an atom or molecule with a photon from the outside.

In the simplest case, which is usually done in monoatomic steam and gas, atom, returning to its initial state, it radiates a photon luminescence identical to the incident (*fig. 14.1*).

$h\nu' = h\nu;$

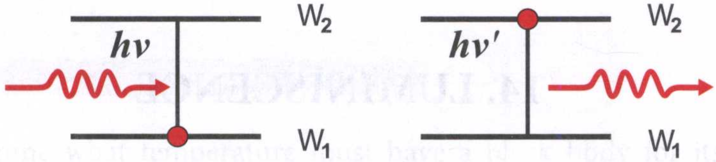


Fig. 14.1.

This phenomenon is called **fluorescence resonance**.

Often, without radiation, atoms from excited states 4-3 pass to level 2, then, spontaneously, to level 1, radiating a photon of lower frequency compared to the frequency of the incident photon (*fig. 14.2*).

Level 2 can belong to an excited atom or another atom. In the first case, luminescence is called **spontaneous**.

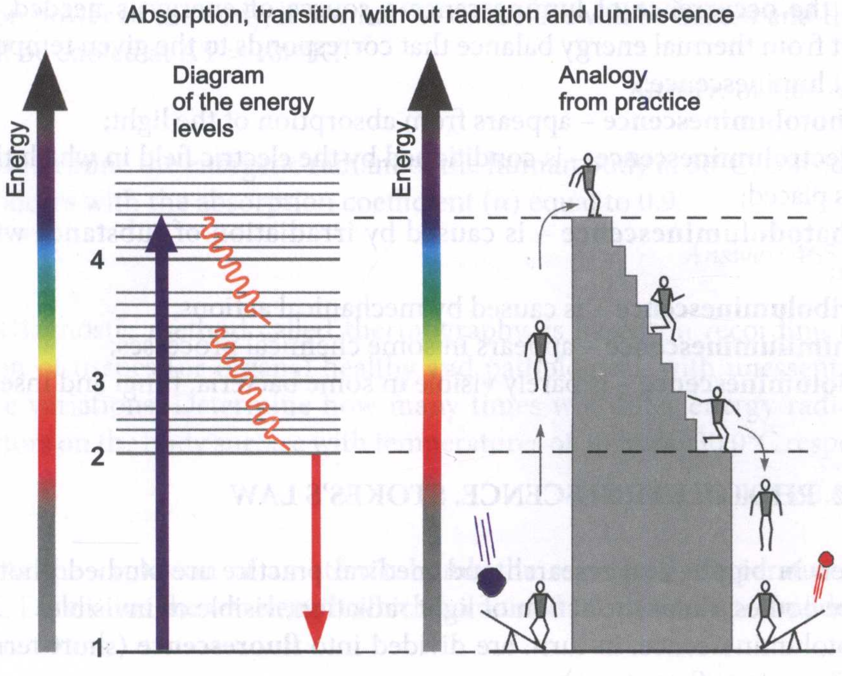


Fig. 14.2.

The second case occurs in gas mixtures whose atoms have the same energy levels.

As an example can serve the laser active medium with **helium** and **neon**, where the excited atoms of helium transfer through collision the energy of the neon atoms, which luminescence represents the **laser radiation**.

In spectroscopy, a **triplet** is considered three spectral lines that are close to each other and determined by the dividing of an energy level of the system into three sublevels. This phenomenon is characteristic for atoms or molecules that have two external electrons. Direct formation of a triplet by absorbing a photon of visible light or ultraviolet is unlikely, because the electron would have to change its orbit and spin simultaneously. For this reason, there is an intermediate state – **singlet**, which usually has a lifetime long enough to allow one of the two electron spin to reverse, so producing a triplet. In this lie one of the mechanisms and the emergence of the phenomenon of **phosphorescence** (*fig. 14.3*).

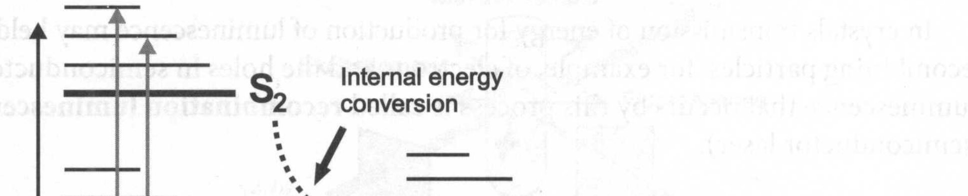


Fig. 14.3.

Usually photoluminescence phenomenon obeys to the **law of Stokes** (1852), which states: **luminescence spectrum relative to the excitation spectrum is shifted to the higher wave length** (*fig. 14.4*).

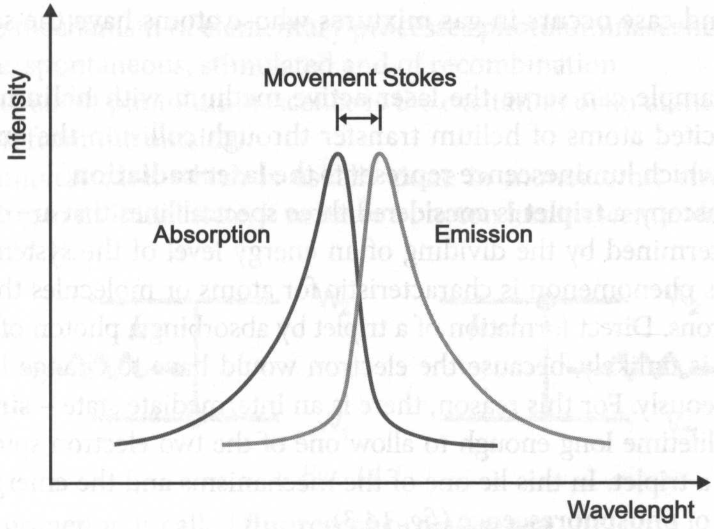


Fig. 14.4.

In crystals transmission of energy for production of luminescence may held by recombining particles: for example, of electrons and the holes in semiconductors. Luminescence that occurs by this process is called **recombination luminescence** (semiconductor laser).

14.3. LUMINESCENCE ANALYSIS (LA)

Research method of different objects based on their luminescence is called **luminescence analysis.**

We investigate the own luminescence of this object, and in case the object is not luminescent, is investigated the luminescence of a **luminophore** where with previously processed.

Luminophores solids and liquids are essential properties of luminescence under the action of various excitatory provided that their energy spectrum to be discrete.

Most of the times, as photoluminescence excitation source are used: **quartz lamp mercury vapor or xenon lamp, laser radiation.**

In some cases the analysis is based on luminescence and other forms of luminescence, such as cathodoluminescence, radioluminescence, chimiluminescence. By this method, after the intensity of spectral lines in the radiation luminescence is determined with a precision of 10^{-4}–10^{-5} g/cm^3 quantitative content of various **substances in pharmaceuticals.**

In favorable conditions allows to detecting luminescence analysis of substances weighing up to 10^{-10}g.

By the luminescence analysis is easily differentiate objects that are identical in appearance (fake currency).

In research, medical and biological luminescent microscope is used extensively in that, unlike optical microscopy, ultraviolet radiation and uses two filters: one is located in front of the condenser and is transparent to ultraviolet radiation only, and the second between objective and eye and only transparent luminescence spectrum.

One relatively simple device used to detect and determine the intensity of luminescent radiation is fluorimeter, a schematic diagram of which is shown in figure 14.5.

LA is widely used in the diagnosis of various diseases of the skin in forensics and many other areas.

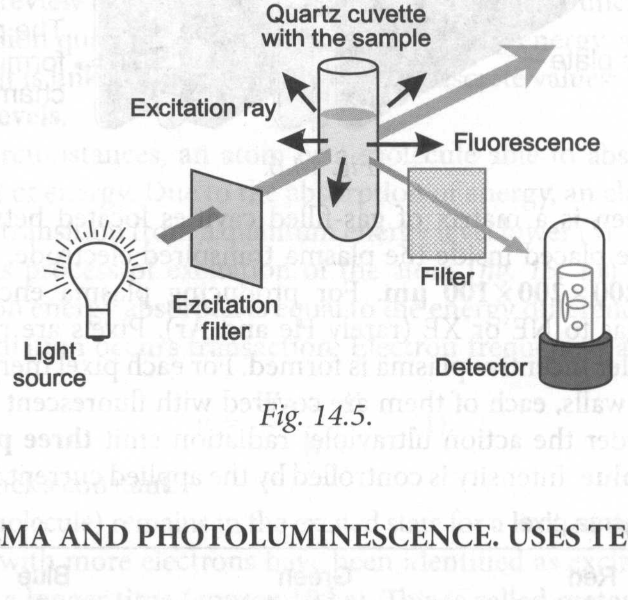


Fig. 14.5.

14.4. PLASMA AND PHOTOLUMINESCENCE. USES TECHNIQUE

As a relative approximation, **plasma** is a physical system consisting of a large number of neutral particles (atoms in the ground state or excited state, photons) and charged particles (ions positive and negative electrons). Whose properties are determined by the collective interactions which macroscopically appear electrically neutral.

Photoluminescence-based products are **gas discharge lamps**, the spectrum of which is close to that of daylight and high efficiency bulbs to incandescence. The most widespread are some tubes contain an inert gas (argon or neon) mercury vapor. While discharging **plasma is formed in this environment that emits**

ultraviolet. The inner surface of the tube is coated with a fluorescent substance which emits a continuous spectrum in visible area. Depending of the gas used to get different colors on emitted radiation (for all red neon, argon blue, yellow sodium vapor).

The construction of televisions, monitors and tablets screen display is used plasma. Two figure 14.6 are the building blocks of a section in a plasma screen.

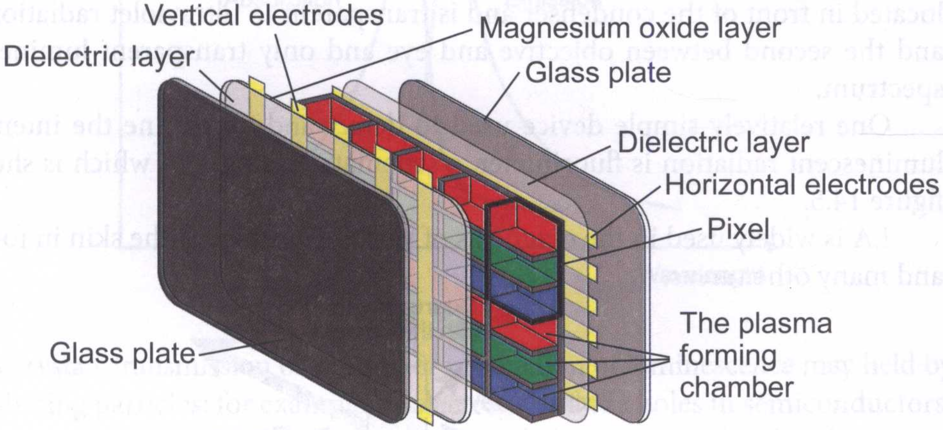


Fig. 14.6.

Plasma screen is a matrix of gas-filled cavities located between two glass plates, which are placed inside the plasma transpired electrode. Pixels have the dimension of **200 × 200 × 100 μm**. For producing plasma enclosures usually are filled with gas to NE or XE (rarely He and Ar). Pixels are rendered with a network of smaller incentive plasma is formed. For each pixel there are three such enclosures. The walls, each of them are covered with fluorescent substances (*fig. 14.7*), which under the action ultraviolet radiation emit **three primary colors: red, green and blue**. Intensity is controlled by the applied current individual cells.

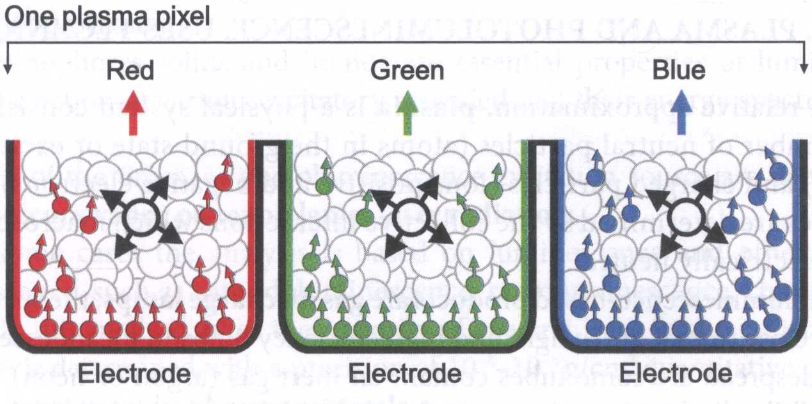


Fig. 14.7.

15. LASERE AND THEIR APPLICATION IN MEDICINE

15.1. THEORETICAL NOTIONS AND PHENOMENA UNDERLYING THE FUNCTIONING OF LASERS

The word "laser" comes from the initials of the English principle of this phenomenon: "Light Amplification by Stimulated Emission of Radiation".

We briefly review the main phenomena underlying the functioning of lasers. Conform emission quantum mechanical laws, electron energy, and therefore the atom to which it is linked, may take only certain discrete values: W_1; W_2; W_3 ... is called **energy levels.**

In some circumstances, an atom or a molecule able to absorb (or emits) a certain amount of energy. Due to the absorption of energy, an electron (or more) may perform a transition from a quantum energy level lower (W_i) a higher energy level (W_S). This process of excitation of the atom (*fig. 15. 1a*) takes place only when the photon energy absorbed is equal to the energy difference between levels of absorbed radiation occurs transaction. Electron frequency (atom) is given by the relation:

$$\nu = \frac{W_S - W_i}{h} \quad (1).$$

Where ***h*** is Planck's constant.

Atom (or molecule) remains in the excited state for a limited time (about 10^{-7} s).

For atoms with more electrons have been identified as excited state electron can remain for a longer time (approx 10^{-3} s). This is called **metastable states** and plays an important role in producing **laser phenomenon**.

In some cases, internal or external, is de-energize the atom and the electron returns to the initial energy, emits a photon whose energy is equal to the photon absorb it.

De-energizing (and excitation) is a random process is conducted at random and statistic. This phenomenon occurs in nature in many ways.

If electron returns by itself, spontaneously, the initial energy level, the phenomenon is called **spontaneous emission** (*fig. 15.1*).

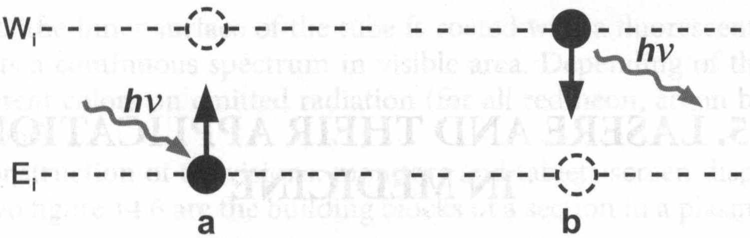


Fig. 15.1.

Another electron, which is in a metastable level, to return to its level W_i, after a long period of time. If after a short time, as a consequence of external causes (for example, electromagnetic radiation), such an electron is imposed to return to level E_i , then the phenomenon is called **stimulated emission** (*fig. 15.2*). In other words, the atom is "spurred" to issue a photon after a much shorter time than would be "life" level respectively if stimulation is excited by a photon whose frequency is that of the radiation equal the excited atom, it is noted that. Beside it (which is not absorbed by the atom) appears another photon (due de-energize atom) of the same energy, direction and phase. If either of these two photons will face each other excited atoms, the same metastable level, and it will stimulate the production of other two emissions of two-photon another (that is, four in total) in step (coherent) with first. If more far, the phenomenon can be repeated, then get photon avalanche multiplications which have the same direction, phase and energy. This energy and amplifying the radiation process is called Emission stimulated **phenomenon (effect) laser**, and the substance of witch atoms allow the production of this phenomenon called **laser active medium**, which may be solid, liquid, gaseous.

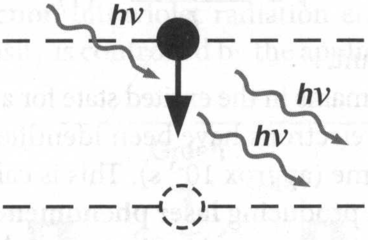


Fig. 15.2.

Population inversion. Typically in atomic systems (quantum), the number of atoms situated on the fundamental (lower energy state of thermodynamic equilibrium) are much larger than the atoms are able excited. In this situation is not possible light amplification because photons emitted by excited atoms are absorbed by non-excited atoms. For laser phenomenon to occur there needs to be a greater number of atoms in the excited state: that it is necessary to reverse the ratio of the number of atoms in the system in favor of that great major excitation. The phenomenon by which the majority of atoms laser active medium is called metastable excited state **population inversion** (*fig. 15.3*).

This inversion involves electron depopulation of lower energy levels of atoms laser active medium and the rise of electron energy levels higher metastable.

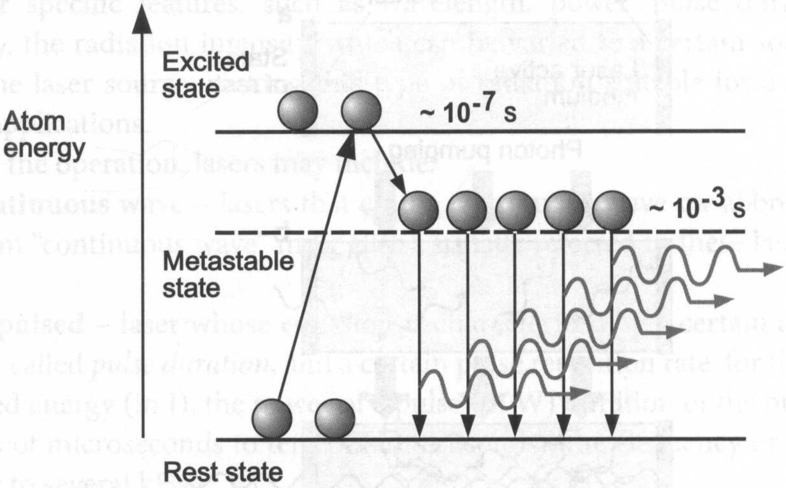


Fig. 15.3. Population inversion.

Photonic pumping (optical). Population inversion is an abnormal situation for an atom because it is the natural tendency of atoms to be in the energy states minim. Meanwhile, this state of thermodynamic equilibrium can not be achieved unless some energy is consumed maintaining them.

The process through which has occurs transmission of the power needed to achieve population inversion is called a **photon pumping** or **optical**. Photon pumping or photon can be done in different ways: photonics, electrical or chemical. Inversion populations can not be obtained by simply heating laser active medium, as at any temperature. Higher energy levels are less “popular” than the inferior energy levels. Source allowing excitation of atoms is called **pumping source**.

Resonant cavity (laser resonator). After obtaining the population inversion, as the initiator of the process of stimulated emission can serve even one atom excited photons emitted by the laser active medium, which in turn stimulates the production of other photons to avoid losing to the outside of the first photon stimulated intermediate and long while their path through the active medium in order de-excitation stimulated a greater number of atoms, it requires a **resonant cavity**.

Resonant cavity photons required to stay a long time in the crowd of excited atoms, thus providing an amplification of such radiation. A resonant cavity may be composed of a system of two mirrors (*fig. 15.4*) rigorously parallel to the role reflect most of the intermediate photon laser. One of mirrors called *extraction mirror* has transparent center, allowing ”exit” laser beam resonant cavity.

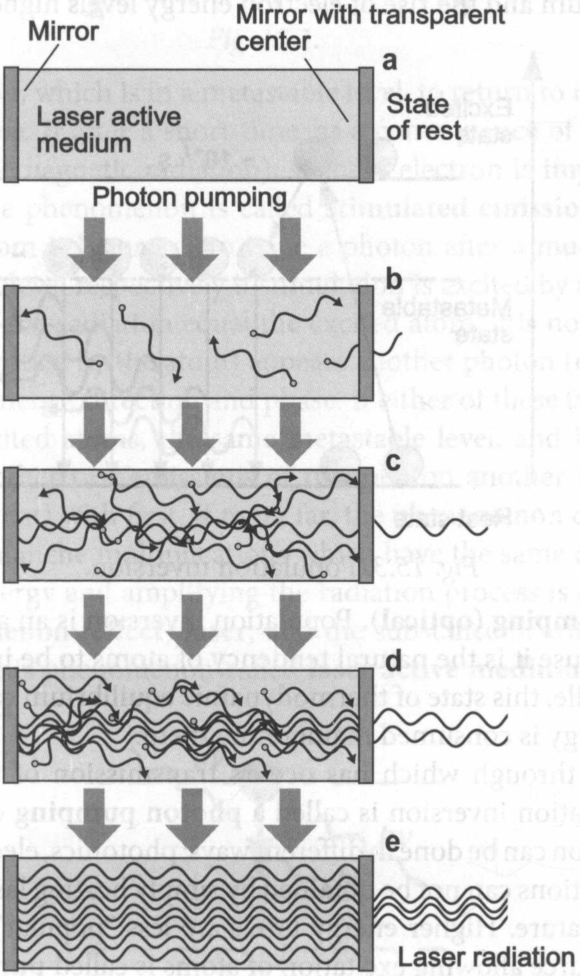


Fig. 15.4.

Whatever the nature of the active medium and pumping system, laser radiation has several general characteristics that distinguish it from traditional sources (lamps):

– **consistency** – the collective property of radiation, this allows interaction conduction packages hence the emergence of interference phenomena and diffraction;

– **monochromaticity** – radiation property laser having a single wavelength photons of all constituents;

– **directionality** – property the laser beam to have a well defined path for each laser beam after a certain distance;

– **brightness** – the property of the laser beam, to have a higher energy density than a conventional light sources.

Other specific features, such as wavelength, power, pulse duration and frequency, the radiation intensity, which can be varied to a certain source or to change the laser source, making this type of radiation suitable for a variety of medical applications.

After the operation, lasers may include:

– **continuous wave** – lasers that emit a continuous wave **cw** abbreviation is used (from "continuous wave" in English), usually referred to these lasers power (in W);

– **in pulsed** – laser whose emission is characterized by a certain amount of emission, called *pulse duration*, and a certain pulse repetition rate, for these lasers mentioned energy (in J), the power of a pulse (in W) duration of the pulse (from hundreds of microseconds to tens of femtoseconds), the frequency or repetition rate (kHz to several kHz).

15.2. ELEMENTS CONSTRUCTING VARIOUS TYPES OF LASER

Laser with He-Ne

To achieve population inversion, often the quality of the laser active medium is used a mixture of two gases .In selecting one of the criteria is that both possess energy levels equal or at least very closer. Photon pumping is performed by electric discharge in a mixture.

The helio-neon laser, role-based gas it is for neon atoms. Diagram gas energy He-Ne mixture is represented in figure 15.5.

The collision energy transfer occurs from atoms helium neon atoms to yield the metastable. Stimulated transitions in the resonant cavity are production the laser fascicle. Mentioning the beam producing the resonant cavity for each gas mixture, which can be used as laser active medium, there is a certain amount of discharge current strength power that ensure the highest population inversion.

Schematic diagram of one He-Ne laser is shown in figure 15.6.

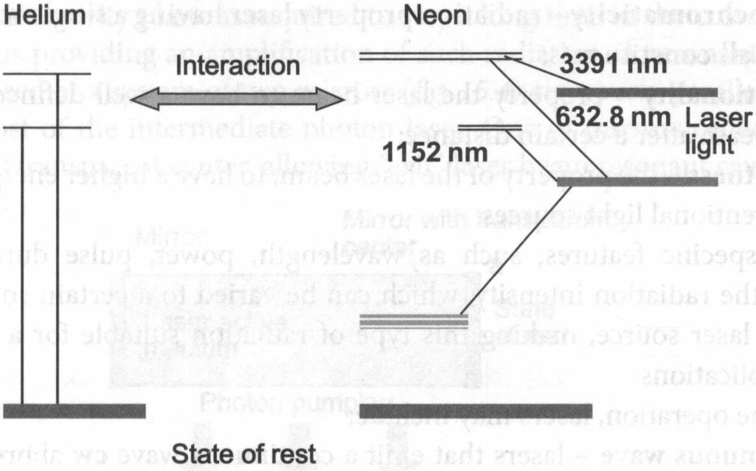


Fig. 15.5.

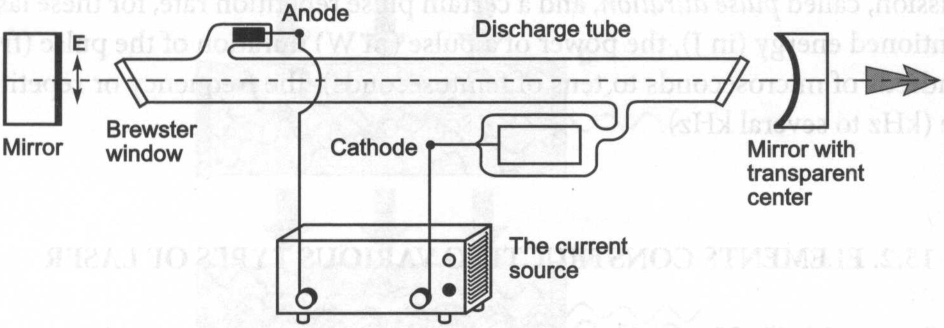


Fig. 15.6.

Brewster window plays an important role in the construction of gas lasers, by:

– avoiding energy loss and exit radiation into electric discharge tube;

– polarizing the laser beam in a plan;

– changing electric discharge tube in case of failure, while maintaining laser resonator mirrors, which are much more expensive.

Laser with CO_2

Active medium is plasma with CO_2, nitrogen and helium for increasing wave efficiency. Most common used length is between 10.6 μm and 9.6 μm, in infrared field. This laser can operate both in continuous and pulsating ways. Is one of the most used lasers in medical practice, in both, excision of tissue, and for vaporization of some injuries. The functional wavelength of this laser is strongly absorbed by water, leading to a small penetration power. This type of laser has many applications in ophthalmology and surgery.

Laser with Ar

The laser with Ar has pressure of 1 torr. Provides radiation with several wavelengths, the most important being 488 nm and 514.5 nm. It operates in continuous way and has special applications in ophthalmology.

Ruby lasers

Ruby laser consists basically of a cylindrical crystal made of ruby, two parallel mirrors, silver or gold and a spiral discharge tube filled with noble gas and connected to a high-capacity capacitor. **Ruby is aluminum oxide, which contains small amounts chromium ions (Cr^{3+} in Al_2O_3)**. Ruby cylinder has only a few cm in length and few mm diameter. From the two mirrors, silver or gold, one of them is compeletly opaque while the other one has a transparent center. They are placed at the end of the cylinder and sometimes can oxidize the ends (*fig. 15.7*). Spiral discharge tube, filled with neon, xenon or neon and krypton mixtures, is connected to a condensor and serves as cameras flashes. This laser operates in pulse mode, and has emission with length of 694.3 nm.

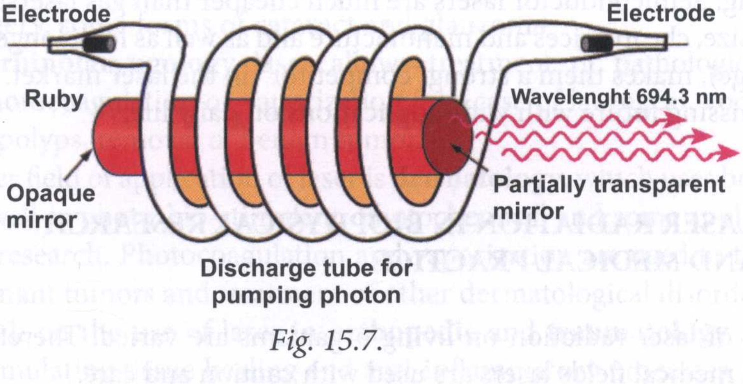


Fig. 15.7.

Semiconductor lasers

Semiconductor laser is formed as the other types of lasers in the same pattern: the active medium, the excitation system, and optic resonator. It is, in fact, a sandwich of three layers made of semiconductor elements with elements of the excitatory system. In this type of laser, the necessary energy to excite the atoms of the active medium and also to inhibit this process is provided by the electric current (*fig. 15.8*). Due to the fact that the sandwich appereance of the laser is similar to the classic diode, from now we will use the term **diode.**

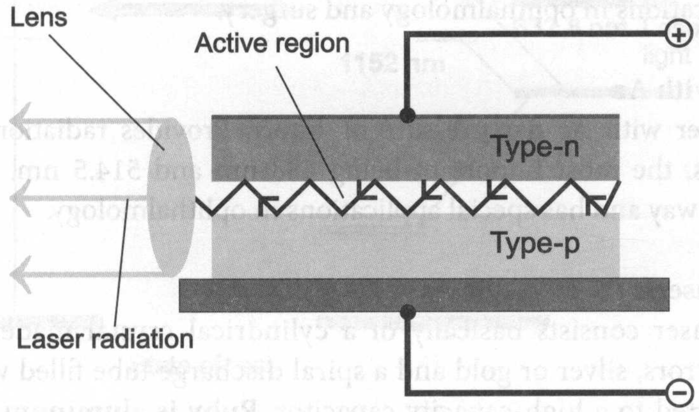


Fig. 15.8.

Diodes laser that are cheap, usable, produced in astronomic quantaties, and very used in electronics.

On the other hand, they are, perhaps, the most fragile devices of laser emission. The fact that the active layer is the size of the bacteria it can easily be destroyed by subjecting it to electrostatic influences, or simply by overheating it. The active layer can destroy itself even without the presence of any of the factors listed above. Simple light emission can vaporize this tiny layer, if the emitted light is too strong. Semiconductor lasers are much cheaper than gas lasers.

Small size, cheap prices and manufacture and as well as long longevity (due to diodes usage), makes them a strong "competitor" in the laser market. Diode laser has a promissing future with wide applications on daily life.

15.3. LASER RADIATION IN BIOPHYSICAL RESEARCH AND MEDICAL PRACTICE

Effects of laser radiation on living organisms are varied. Therefore, in biological and medical fields lasers are used with caution and care.

As shown in specialized research, when interacting with living mater, laser radiations may produce the following effects: the thermal, photo-biostimulent and photochemical. According to the temperature reached when in contact with the material, two types of thermal effects are observed: **coagulation** (for temperatures between 60 and 100°C) and **volatilization** (for temperatures higher than 100°C).

Since the beginning of its appearance, the laser has found numerous applications in different new areas of activity (metrology, information processing and transmission, holography, cybernetics and so on). Moreover, there is a more significant increase of laser applications in medical and biological fields.

In medicine, the introduction of laser medical techniques allows the development of more efficient techniques to replace the older ones or creates new ways of investigation and treatment. So with a laser device emitting infrared radiation, it becomes possible to determine the substances in the blood without the need of collecting it. This device also allows quick and accurate measurement of blood glucose, fats (cholesterol), uric acid and alcohol (ethanol), which aid in early detection of several ilnesses (**tuberculosis, diabetes** etc.).

In **surgery** and **microsurgery**, laser is used in treatment of gliomas, separation of tumors from the main blood vessels, vaporization of tumors from the ventricles, excision of an intense cerebral vascular tumor, anostomosis (fusion) of blood vessels, deatachmente of the muscles from the bone, clinical endoscopy for controlling gastrointestinal bleeding, for cicatrization in ulcers, etc.

In the new surgical techniques, the laser beam serves as a scalpel some negative effects on adjacent tissues can be avoided. The bid advatages of these new surgical procedures don't have complications, and hemorrhages don't occur as frequent.

In **ophthalmology**, the laser is used in: diabetic retinopathy, retinal vessel occlusion, prevention and extension of bleeding or tearing of the retina, eyelid tumor surgery, some forms of cataract and glaucoma.

In **otorhinolaryngology** laser allows treatment of pathological damage through photocoagulation or vaporization – in case of laryngeal, vocal nodules, vocal cord polyps, removal of benign tumor.

Another field of application of laser is **dermatology**, which uses both thermal effects, as well as photo-bio-stimulant. Photochemical and some applications are in current research. Photocoagulation and vaporization are used to treat couperose, malignant tumors and treatment of other dermatological disorders.

Research on the use of laser in **orthopedic** and **traumatology** showed the effect of stimulating tissue healing and anti-inflammatory processes. As a result, laser radiation is now indicated for the treatment of a whole series of problems: post-traumatic arthrosis and rheumatism, periarthritis limestone, osteoporosis, consolidations delay, etc. discartoze and spondylosis.

Therapy with the help of laser radiation involves a wide range of laser devices with optimal and functional characteristics for the type of illness. From the all types of laser, CO_2 laser, He-Ne laser Ar and others are the most used in medicine.

It is obvious that in some cases, laser therapeutical techniques can't replace the classic ones, but they aid in precision, efficiency and speed.

EXERCISES

1. What will be the length of space radiation of an pulse of laser radiation with duration of 10^{-12} s?

Answer: 0.3 mm.

2. Using ruby laser ($\lambda = 694.3$ nm) and a diffraction network, maximum of order 1 per screen located at a distance of 0.5 m is moved in relation to the maximal cross 3 cm. Determination constant of diffraction grating.

Answer: ≈ 0.012 mm.

3. Wavelength of laser radiation is 750 nm. Light beam from this laser light an area of 1 mm², knowing that the laser power is 200 mW, determine the number of photons that bombard the surface within 10 seconds.

Answer: $57 \cdot 10^{16}$ photons.

4. Two lasers emit radiation at $\lambda_1 = 632.8$ nm and $\lambda_2 = 514.5$ nm respectively. Determine how much the radiation energy of these photons will differ.

Answer: with 0.44 eV.

5. An output of 200 mW laser emits radiation with a wavelength equal to 570 nm. Determine the number of photons emitted by this laser during 10 seconds.

Answer: $57 \cdot 10^{17}$.

16. MAGNETIC PROPERTIES OF SUBSTANCES. MAGNETIC RESONANCE

16.1. MAGNETIC FIELD. BASIC CONCEPTS

A magnetic field represents a form of matter through which the action of force on moving electric charges occurs.

Quantitative characteristic of a magnetic field in vacuum is called **the intensity $\vec{H}$** and in an unlimited homogeneous material medium – **magnetic induction $\vec{B}$**. Between these quantities the following relationship occurs:

$$\vec{B} = \mu_0 \mu \vec{H} \quad (1).$$

Where: μ_0 – magnetic constant (**$4\pi \cdot 10^{-7}$ N/m**); μ – relative magnetic permeability of the environment; $\mu_a = \mu_0\mu$ – the absolute magnetic permeability of the environment.

The magnetic and electrostatic fields are graphically represented using force lines called *magnetic induction lines* (*Fig. 16.1*). These lines are drawn so that the tangent to them at any point should coincide with the direction vector $\vec{B}$. Unlike electrostatic field lines, magnetic induction lines are closed.

The number of magnetic induction lines, which crosses a unit area perpendicular to the direction of the field, coincides with the module of vector $\vec{B}$.

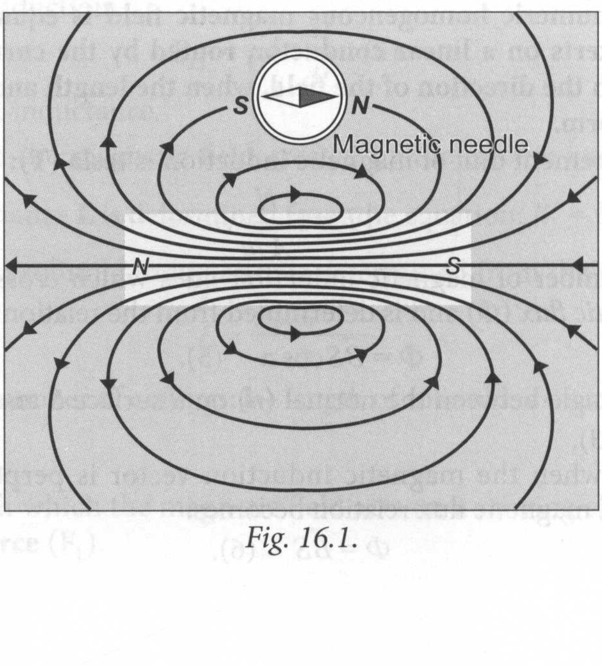


Fig. 16.1.

The Ampere's law determines the force used by a uniform magnetic field ($\vec{B}$ – constant, induction lines are parallel and equidistant) on a linear conductor crossed by electricity (*fig. 16.2*):

$$\vec{F} = I\vec{B}\,l\sin\alpha \quad (2).$$

Where: ***I*** – the current intensity; ***α*** – the angle between vector $\vec{B}$ direction and the conductor; ***l*** – the conductor length.

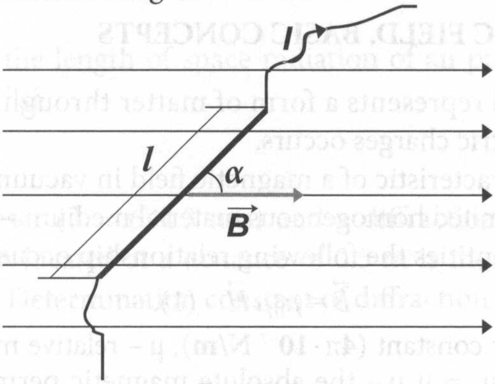


Fig. 16.2.

When α = 90°, the force (F) has the maximum value:

$$\vec{F} = I\vec{B}\,l \quad (3) \text{ and}$$

$$\vec{B} = \frac{\vec{F}_{\max}}{I\,l} \quad (4).$$

The last relationship allows formulating the physical meaning of the magnetic induction:

induction of a numeric homogeneous magnetic field is equal with the force that this field exerts on a linear conductor, routed by the current and located perpendicular on the direction of the field, when the length and intensity of the current are uniform.

In SI, measurement unit of magnetic induction is **tesla (T)**:

$$1T = 1\frac{N}{A\,m}.$$

The total number of magnetic induction lines which cross a surface (*S*) is called *the magnetic flux* (*Φ*) and is determined from the relation:

$$\Phi = \vec{B}S\cos\alpha \quad (5).$$

Where: *α* – the angle between the normal (***n***) on a surface ***S*** and the direction of vector $\vec{B}$ (*fig. 16.3*).

In the case when the magnetic induction vector is perpendicular to the surface ***S*** (***α*** = **0**), magnetic flux relation becomes:

$$\Phi = \vec{B}S \quad (6).$$

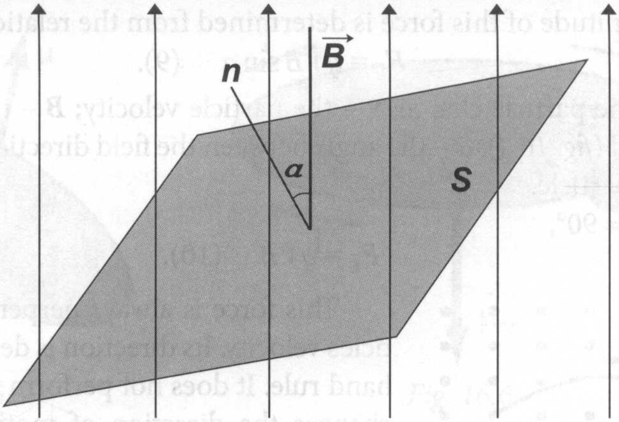


Fig. 16.3.

The measurement unit of the magnetic flux in SI is called **Weber (Wb)**:

$$1\ \text{Wb} = 1\ \text{T m}^2.$$

At the variation of the magnetic flux passing through a closed loop, an electromotive force of electromagnetic induction, proportional to the speed of magnetic flux, always occurs.

M. Faraday discovered **the fundamental law of electromagnetic induction** in 1831. Mathematical expression of this law is given by:

$$E_i = -\frac{\Delta\Phi}{\Delta t} \quad (7).$$

The appearance of an induced electromagnetic field on a circuit when there is a variation on the current intensity is called **self-induction** – a particular case of electromagnetic induction:

$$E_{s.i.} = -L\frac{\Delta I}{\Delta t} \quad (8).$$

Where: L – circuit inductance.

The energy of the magnetic field created by the electric current through a circuit with inductance **L** is determined from the equation: $W = L\frac{I^2}{2}$.

In particular a solinoid inductance is determined from the relation:

$$L = \frac{\mu_0\,\mu_r\,n^2\,s}{l}.$$

Where: **n** – total number of coilings; **l** – length of solinoid; **s** – sectional area.

Lorentz force

The force with which the magnetic field acts on a moving electric charge is called **Lorentz force** (F_L).

The magnitude of this force is determined from the relation:

$$F_L = q v B \sin\alpha \quad (9).$$

Where: **q** – the particle charge; **v** – the particle velocity; **B** – the induction of the magnetic field (*fig. 16.4*); **α** – the angle between the field direction and the direction of a moving particle.

When $\alpha = 90°$,

$$F_L = q v B \quad (10).$$

This force is always perpendicular to the particle's velocity. Its direction is determined by the left hand rule. It does not perform work, but gradually changes the direction of motion of the particle. The radius of the curvature of the trajectory can be easily determined. Simultaneously, the particle is subjected to the action of two opposite forces:

– The centripetal magnetic force:

$$F_L = q v B.$$

– The mechanical centrifugal force:

$$F' = \frac{m v^2}{r}.$$

Where: **r** – the radius of the curvature of the trajectory; **m** – the mass of the particle.

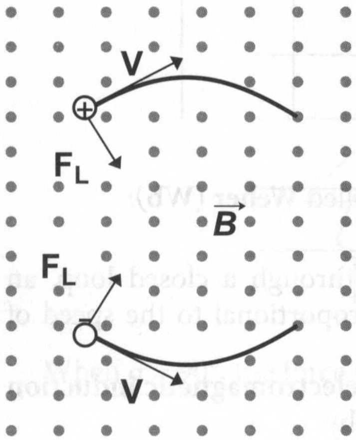


Fig. 16.4.

Since these forces are equal in module, we obtain:

$$q v B = \frac{m v^2}{r} \Rightarrow r = \frac{m v}{q B} \quad (11).$$

16.2. MAGNETIC MOMENTS OF THE ELECTRON

It is known that electrical charges in movement (electric current) create a magnetic field. By definition, the magnetic moment (**μ**) of a circular electric curent is considered to be the product of current intensity (**I**) and the surface (**S**) limited by the circuit (*fig. 16.5*):

$$\vec{\mu} = IS = I\pi r^2 \quad (12).$$

The direction of this moment is determined according to the rule of drill.

A magnetic moment is created and by the microcurrent (**i**) resulting from the movement of electrons around the nucleus is called **the orbital magnetic moment** (*fig. 16.6*):

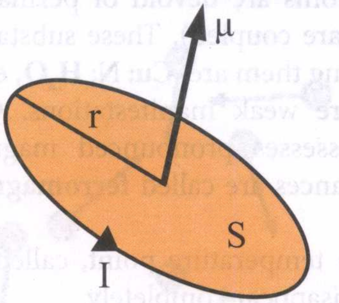


Fig. 16.5.

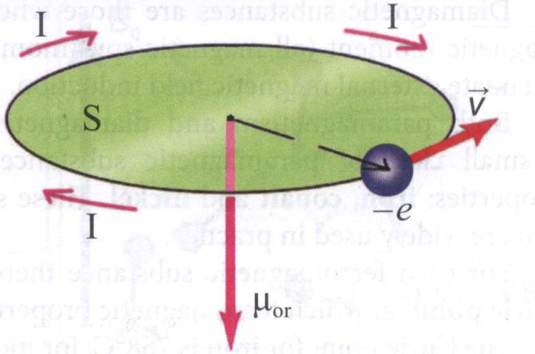


Fig. 16.6.

$$\vec{\mu} = i\pi r^2 = \frac{eV}{2\pi r}\pi r^2 = \frac{eVr}{2} \quad (13).$$

Where: *V* – velocity of the electron; *r* – radius of the orbit.

It was found that the electron performs also a rotational motion around its axis called **the spin movement**. This movement represents an electric charge in motion and, a microcurrent which gives to the electron a second magnetic moment is called **spin magnetic moment** (μ_S).

16.3. PARAMAGNETIC AND DIAMAGNETIC SUBSTANCES

According to the magnetic properties all substances in nature are divided in two classes: **paramagnetic and diamagnetic substances.**

The primary role of this division belongs to electron magnetic spin moments.

Paramagnetic substances are those, which have atoms that are in permanent magnetic moment because a part of the magnetic spin moments of the electrons within the atoms remain uncompensated (not all magnetic spin moments are coupled) (*fig. 16.7*).

These substances enhance the external magnetic field. For example, among them are: **Pb**; **O_2**; **Al**, etc.

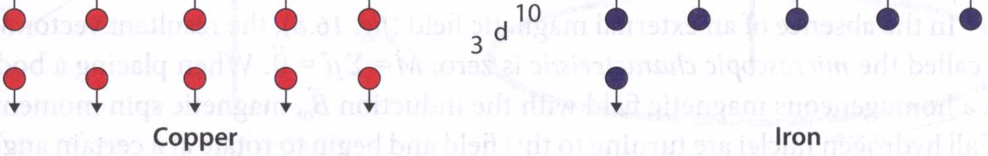


Fig. 16.7.

Diamagnetic substances are those whose atoms are devoid of permanent magnetic moment (all magnetic spin moments are coupled). These substances attenuate external magnetic field induction. Among them are: **Cu**; **N**; **H_2O**, etc.

Both paramagnetism and diamagnetism are weak manifestations. Only a small class of paramagnetic substances possesses pronounced magnetic properties: **iron, cobalt** and **nickel**. These substances are called **ferromagnetic** and are widely used in practice.

For each ferromagnetic substance there is a temperature point, called **the Curie point**, at which ferromagnetic properties disappear completely.

The Curie point for iron is 768°C; for nickel – 358°C and for cobalt – 1000°C.

Largely biological tissues, as well as water, are diamagnetic. In the human body only some molecules and molecular aggregates possess paramagnetic properties. One **microorganism**, called **Spirella**, synthesize **ferritin**, accumulating it in special organs called **magnetosome**.

Magnetosomes can be considered magnetic needles that provide a directional orientation in space using the Earth's magnetic field. It is assumed that bees, butterflies, birds and dolphins have such magnetic "guides".

Humans also have a ferromagnetic substance concentration in the **adrenal glands**; its function it is not determined even until now.

16.4. MAGNETIC RESONANCE IMAGING (MRI)

Components of atomic nuclei – protons and neutrons also have magnetic moments. The result of these moments is the magnetic spin moment of the nucleus. This moment is different from zero, when the nucleus contains an odd number of protons and neutrons. As a result, a special interest is taken for hydrogen atoms, which have a single proton in the nucleus, which allows the hydrogen atoms to possess permanent magnetic spin moment and is the most sensitive to an external magnetic field.

On the other hand, hydrogen atoms are found most frequently in the human body (63%).

In a body the hydrogen atoms (either biological tissue) magnetic moments $\vec{\mu}$ characteristic to proton are oriented in all directions.

In the absence of an external magnetic field (*fig. 16.8*), the resultant vector $\vec{M}$, is called the *microscopic characteristic* is zero: $\vec{M} = \Sigma\vec{\mu} = \vec{0}$. When placing a body in a homogeneous magnetic field with the induction $\vec{B}_0$, magnetic spin moments of all hydrogen nuclei are turning to this field and begin to rotate at a certain angle around the axis which coincides with the direction of the induced field $\vec{B}_0$.

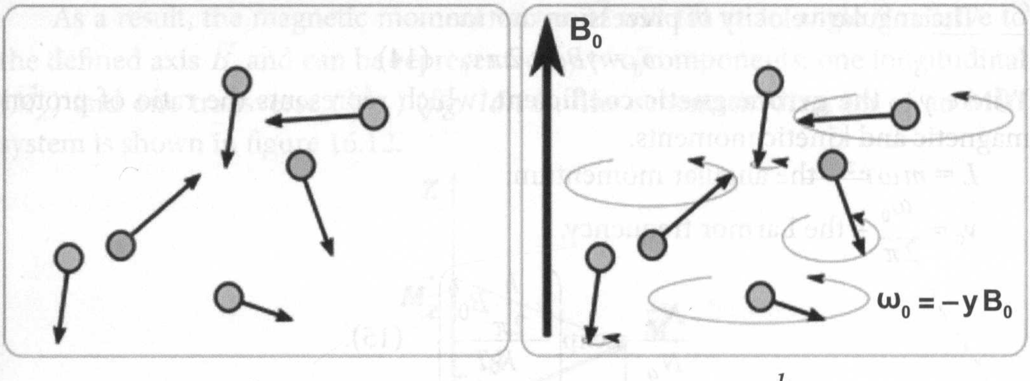


Fig. 16.8.

This motion is called the *precision motion* – the proton axis is described as a cone. The precision motion of a proton's axis, under the action of magnetic field, separates protons in two categories:

1) a greater number with spin $+\frac{1}{2}$ aligned in the direction of the field $\vec{B}_0$ (*fig. 16.9 a*);

2) a smaller number with spin $-\frac{1}{2}$ aligned in the opposite field direction $\vec{B}_0$ (*fig. 16.9 b*).

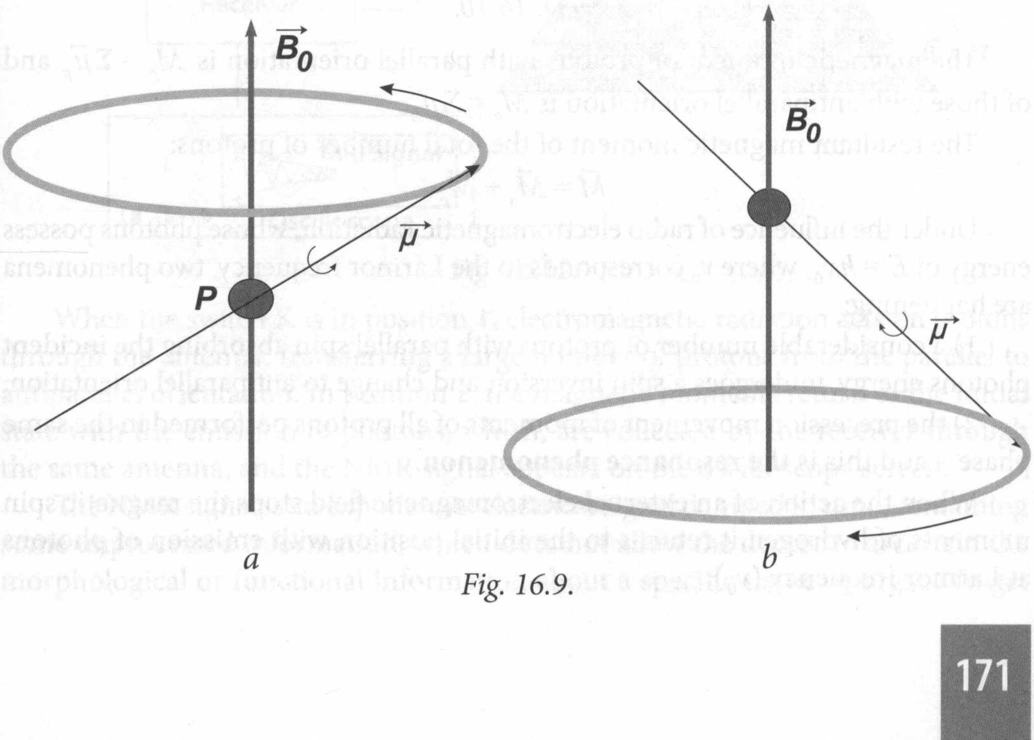


Fig. 16.9.

The angular velocity of precession motion:

$$\vec{\omega_0} = \gamma B_0 = 2\pi \nu_0 \quad (14).$$

Where γ – **the gyromagnetic coefficient**, which represents the ratio of proton magnetic and kinetic moments.

$L = m\omega r^2$ – the angular momentum;

$\nu_0 = \dfrac{\omega_0}{2\pi}$ – the Larmor frequency.

$$\frac{N_p}{N_a} = \exp\left(\frac{\gamma \dfrac{h}{2\pi} B_0}{KT}\right) \quad (15).$$

Where: N_p – number of protons with "parallel" spin; N_a – number of protons with "antiparallel" spin; K – Boltzmann constant; T – absolute temperature.

External magnetic field $\vec{B_0}$ distributes protons in two energy sublevels (*fig. 16.10*).

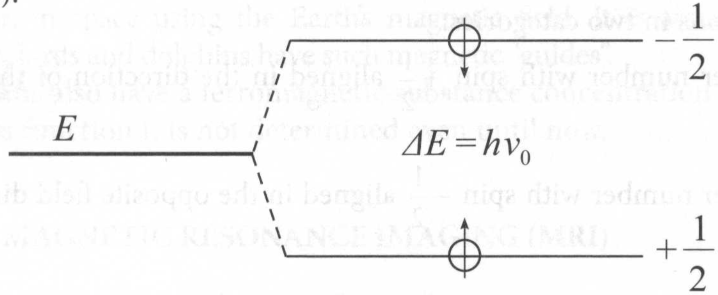


Fig. 16.10.

The magnetic moment of protons with parallel orientation is $\vec{M}_p = \Sigma\vec{\mu}_p$ and of those with antiparallel orientation is $\vec{M}_a = \Sigma\vec{\mu}_a$.

The resultant magnetic moment of the total number of protons:

$$\vec{M} = \vec{M}_a + \vec{M}_p.$$

Under the influence of radio electromagnetic radiation, whose photons possess energy of $E = h\nu_0$, where ν_0 corresponds to the Larmor frequency, two phenomena are happening:

1) a considerable number of protons with parallel spin absorbing the incident photons energy, undergoes a spin inversion and change to antiparallel orientation;

2) the precession movement of moments of all protons performed in the same phase – and this is **the resonance phenomenon**.

When the action of an external electromagnetic field stops the magnetic spin moments of hydrogen it returns to the initial position with emission of photons at Larmor frequency (ν_0).

As a result, the magnetic moment vector $\vec{M}$ will tilt at an angle θ relative to the defined axis $\vec{B}_0$ and can be represented by two components: one longitudinal ($\vec{M}_Z$) and one transverse ($\vec{M}_{XY}$) (*fig. 16.11*). The schematic diagram of an **MRI** system is shown in figure 16.12.

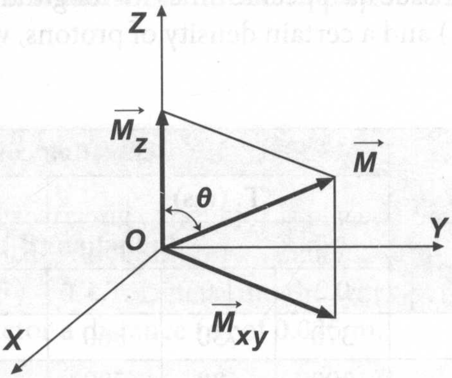


Fig. 16.11.

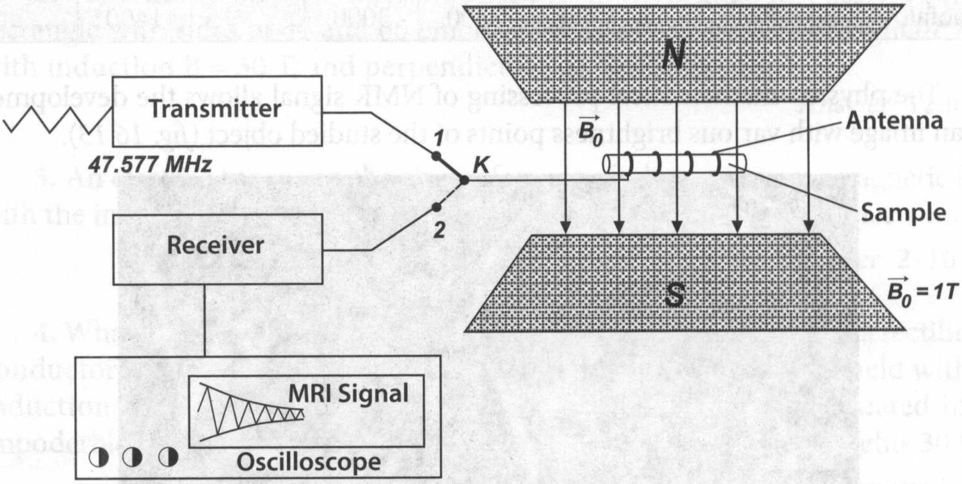


Fig. 16.12.

When the switch K is in position 1, electromagnetic radiation acts on protons through the antenna, transferring a large number of protons from the parallel to antiparallel orientation. In position 2, the magnetic moments return to the initial state with the emission of photons, which, are collected by the receiver through the same antenna, and the NMR signal appears on the oscilloscope screen.

The NMR signal is an exponential electrical signal damped in time, containing some unprocessed information, which does not allow the doctor to ascertain the morphological or functional information about a specific tissue or organ. To get the desired visual information special devices including a high-speed computer, are used in the installation diagram.

By nuclear spin relaxation the dynamic process (developed in time) of recovery for the system and its components to the initial state of equilibrium is inferred.

Each substance or tissue has specific times for longitudinal (T_1) and transverse relaxation (T_2) (*table 1*) and a certain density of protons, which is reflected in the NMR signal.

Table 1

Biological substance	**Relaxation constants**					
	T_1 (ms)			**T_2 (ms)**		
	Induction (B_0) (tesla)					
	0.2	1.0	1.5	0.2	1.0	1.5
Muscle	370	730	860		45	
Nerve tissue gray	390	680	780		90	
White matter nerve tissue	490	810	920		100	
Chefalo-rahidian fluid	1400	2500	3000		1400	

The physico-mathematical processing of NMR signal allows the development of an image with various brightness points of the studied object (*fig. 16.13*).

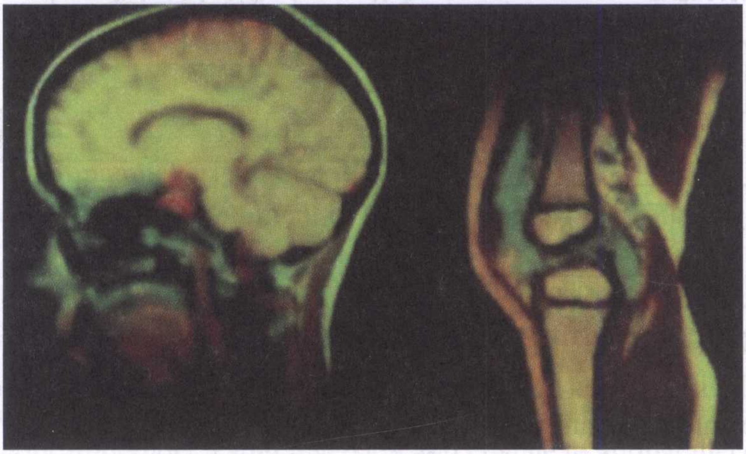


Fig. 16.13. NMR images.

The Nuclear Magnetic Resonance as a phenomenon was discovered in 1946. Shortly after NMR was introduced in the diagnostic arsenal of modern medicine, the term "nuclear" was excluded in order not to produce any association with the nuclear weapon.

The **magnetic resonance tomography (MRT)** is currently used in the world of medical literature. This is the most modern method and it lacks any harmful effects in medical diagnostics.

The importance of NMR in contemporary science is reflected in three Nobel Prizes with which it was mentioned: two in physics (1952, 1991) and one in medicine.

EXERCISES

1. A linear current-carrying conductor is placed in a homogeneous magnetic field perpendicular to the induction lines. Cable length (l) – 0.2 m, current (I) – 8 A, induction field ($\vec{B}$) – 0.4 T. Determine the work performed in ampere force for moving the conductor a distance (d) of 0.025 m.

Answer: $16 \cdot 10^{-3}$ J.

2. How many lines of conventional magnetic induction pass through a rectangle with sides of 40 and 60 cm, located in a homogeneous magnetic field with induction B = 50 T, and perpendicular on field direction?

Answer: 12 lines.

3. An electron moves on the circumference of a homogeneous magnetic field with the induction B = 0.18 T. Determine the rotation period of the electron.

Answer: $2 \cdot 10^{-10}$ s.

4. What intensity should an electric current have passing through a rectilinear conductor? The conductor is placed in a homogeneous magnetic field with an induction of (B) – 20 T perpendicular to the induction lines, situated in an impoderble state. It is known that 1 m in length of the conductor weights 30 N?

Answer: 1.5 A.

5. An electron enters a magnetic field with an induction (B) of 0.05 T, in the direction perpendicular to the induction lines, with a speed of (v) de $4 \cdot 10^4$ km·s^{-1}. Determine electron trajectory radius of curvature.

Answer: 4.55 mm.

6. What is the magnetic moment of a solenoid with 500 turns, and a current intensity of 0.3 A, if the surface area of one turn is 200 cm^2?

Answer: 3 A·m^2.

7. Calculate hydrogen gyromagnetic constant which resonates at 42.57 MHz in a homogeneous field with the induction (B_0) equal to 1 T.

Answer: $\gamma = 267.47 \cdot 10^6$ (S.I.).

8. The ratio between the gyromagnetic constant of ^{19}F and of hydrogen is 0.94. Determine the resonance frequency of ^{19}F in a field with induction (B_0) equal to 1 T. It is known that in this field for hydrogen $\gamma_H = 267.47 \cdot 10^6$ (SI).

Answer: $\nu_0 = 40.03$ MHz.

9. According to Bohr's theory, in a non-excited hydrogen atom the electron moves around nucleus in a circumference with a radius of $0.53 \cdot 10^{-8}$ cm and with a speed of $2.2 \cdot 10^6$ cm/s. Determine the orbital magnetic moment of the hydrogen atom.

Answer: $0.93 \cdot 10^{-25}$ $A \cdot m^2$.

10. A conductor with a length of 6 m is moving perpendicular to a homogeneous magnetic field with the induction of 1.2 T and with the speed of 15 m/s. Determine the e.m.f. induction that appear at the ends of the conductor.

Answer: 108 V.

11. What induction should a coil have, in order to obtain the magnetic field energy equal to 1 J, when the electricity passeing through it has an intensity of 2 A?

Answer: 0.5 H.

12. We have a coil with the inductance of 0.01 H and a current with the intensity of 20 A. Determine the e.m.f. of self-induction, which appears along the interruption of the circuit, for a period of $2 \cdot 10^{-3}$ s.

Answer: –100 V.

13. A proton and an electron, accelerated by the same potential difference, enter a uniform magnetic field. What particle has a greater trajectory radius and by how many times?

Answer: protons; 43 times.

17. THE STABILITY OF THE ATOMIC NUCLEI

17.1. THE NUCLEAR FORCES. THE MASS DEFECT AND THE BINDING ENERGY

The electrical forces are sufficient to explain the connection between electrons in an atom, but they cannot explain the way positively charged protons and neutrons, that do not have electric charge, form stable systems. The connection between protons and neutrons in an nucleus must be attributed to the other types of forces – **the nuclear forces.**

The electrical forces act at long distances, while the nuclear forces act at very small distances (**$r \approx 10^{-15}$ m**) (for comparison: the radius of the nucleus is approximately 10^{-14} m). Also, these forces act between all nucleons, imprevious to the fact that they do or do not have an electric charge. Nuclear forces are stronge than electric overcoming the opposing forces between protons.

The mass defect. It was determined that the mass of all the atomic nuclei is smaller than the masses of protons and neutrons, that it forms in free state.

The magnitude of this difference of mass is called **the mass defect** and it's value is determined by the formula:

$$\Delta \mathbf{m} = \mathbf{Zm_p} + (\mathbf{A} - \mathbf{Z})\mathbf{m_n} - \mathbf{M_N} \quad (1).$$

Where: A – the atom's mass number; Z – the running number; m_p – the proton's mass; m_n – the neutron's mass; M_N – mass of the nucleus.

The binding energy. According to Einsten's relation concerning equivalency between mass and energy, the defect of mass has a particular quanity called **the binding energy (W_{bind}):**

$$\mathbf{W_{bind}} = \Delta \mathbf{m}\,\mathbf{c}^2 \quad (2).$$

Where: Δm – the mass defect of nucleus; c – the speed of light.

In other words, the binding energy is the mechanical work which must be performed in order to separate one atomic nucleus into individual nucleons.

The most suitable magnitude for studying the stability of nuclei is **the binding energy of a nucleon (B)**, defined as the binding energy of a nucleus divided into the number of nucleons (*fig. 17.1*).

$$B = \frac{W_{lg}}{A} \quad (3).$$

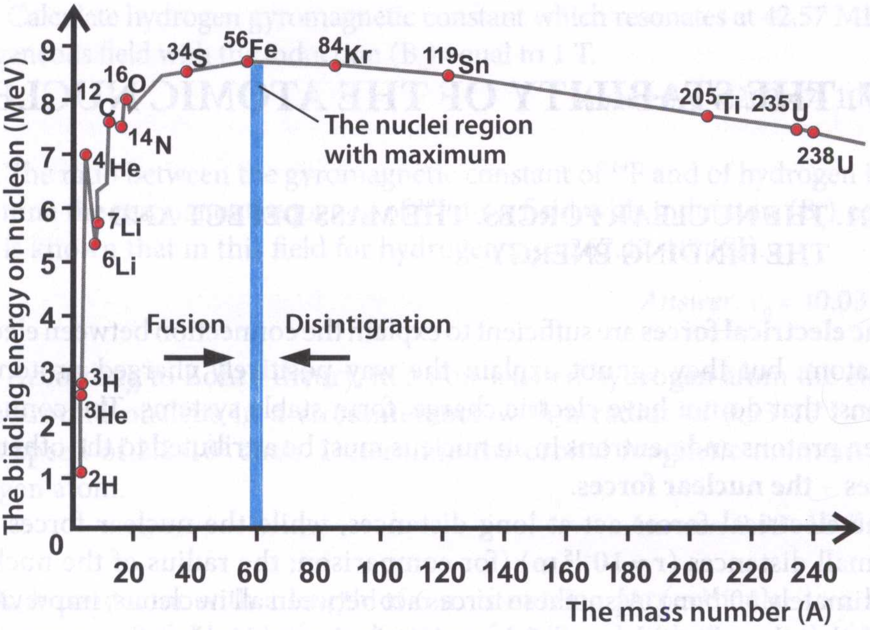


Fig. 17.1.

A high value of the binding energy of a nucleon means a high stability of the nucleus.

Figure 17.1 demonstrates that nuclei with **A**, located between 40 and 140, have B ≈ 8.5 MeV. The heavy nucleons B decrease to ≈ 7.5 MeV.

The process whereby the heavy nuclei increase their B is **the α disintegration:**

$$^{210}_{84}\text{Po} \rightarrow {}^{206}_{82}\text{Pb} + {}^{4}_{2}\text{He} \quad (4).$$

For the light nuclei B increases to A=40. Moreover, there exists pronounced variations for B. Nuclei A = 4; 8; 12; 16 have the B bigger beside of neighboring nuclei. This indicate that groups of 4 nucleons (2 protons and 2 neutrons) are very stable.

Increasing of the **B** in the crossing from light nuclei to heavy nuclei is possible through a process of uniting the light nuclei, called **fusion** (*fig. 17.2*):

$$^{2}_{1}\text{H} + {}^{3}_{1}\text{H} = {}^{4}_{2}\text{He} + {}^{1}_{0}\text{n} \quad (5).$$

The processes of fusion is possible at very high temperatures. The thermonuclear reactions represent **source of solar energy**.

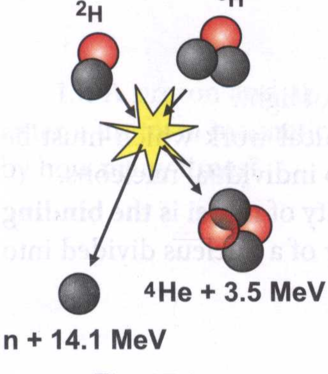


Fig. 17.2.

17.2. THE RADIOACTIVITY

Radioactivity (lat. ***radius*** – *radius, radiation*) is a phenomenon that results after the radioactive decay of atoms or of their nuclei. It is the process whereby the nucleus of an atom are spontaneously transformed into an another atomic species. The element which nuclei can suffer a radioactive decay is called a **radioactive isotope**. Usually, the transformation is accompanied by the expulsion of some subatomic particles, having a very big speed as the emission of electromagnetic radiation is accompained. Radioactivity is an exothermic phenomenon (the release of energy to the environment).

During radioactive disintegration, two types of "radiation" is emitted:

1) the subatomic particles (initially they were called *rays*, because their nature was not known):

- nuclei of the Helium (He^{2+}) that have a high speed, called the *α rays*;
- electrons, called the *β rays*;
- positrons, called and β^+ *rays*;
- neutrons;

2) the electromagnetic waves of a high energy (high frequency or, equivalent, small wave length), also called *gama (γ) radiation (rays)*.

All these "radiations" have the property to ionize the environment through the medium that they pass, thus making them an electrical conductor. This is why these "radiations" are called **the ionized radiations.**

17.3. THE NUCLEAR TRANSFORMATIONS. THE TYPES OF THE NUCLEAR RADIATION

In most of the cases of radioactive decay, there is a transformation of one nucleus into the nucleus of another atom.

The α disintegration (*fig. 17.3*) produces one nucleus of an atom with the atomic number less by 2 and the mass number less by 4, according to the general expression of the reaction:

$$ {}_{Z}^{A}X \rightarrow {}_{Z-2}^{A-4}Y + {}_{2}^{4}\alpha \qquad (6). $$

The energy of particles **α** is well determined. This is why their emission forms a linear spectrum, energies being contained between 4 MeV and 10 MeV.

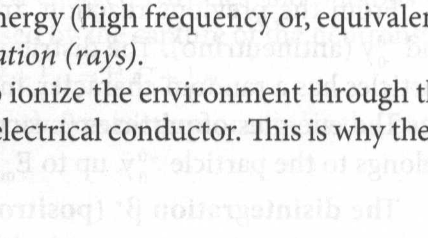


Fig. 17.3.

The β disintegration (*fig. 17.4*) produces one nucleus with the atomic number higher by 1 and with the same mass number. During β half-life occurs a transformation of one neutron into one proton and one electron, the last being expelled as the β half-life. The general expression of the reaction is:

$$^{A}_{Z}X \rightarrow {}^{A}_{Z+1}Y + {}^{0}_{-1}e + {}^{-0}_{0}y \quad (7).$$

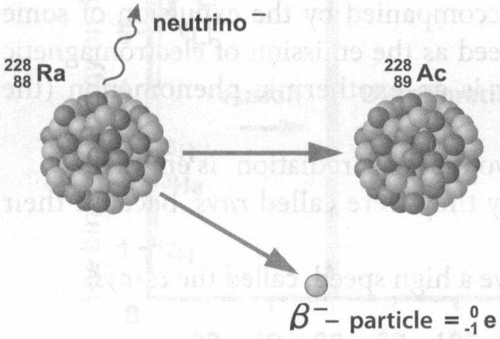


Fig. 17.4.

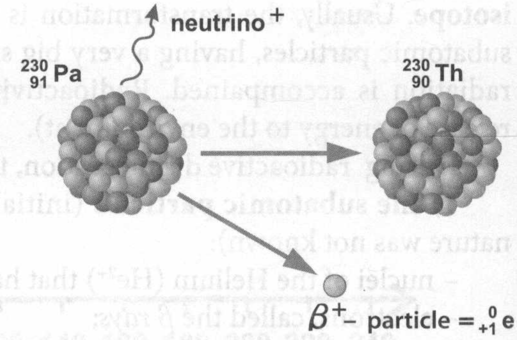


Fig. 17.5.

From the previous reaction, from nucleus are excluded two particles – $^{0}_{-1}$e and $^{-0}_{0}$y (antineutrino). The distribution of the emitted energy between these two particles has a random character, making the β emission spectrum continuous.

The energies of particles β vary from 0, when all the energy of the reaction belongs to the particle $^{-0}_{0}$y, up to E_{max}, when $^{-0}_{0}$y lacks of energy.

The disintegration β⁺ (positrons) (*fig. 17.5*) is produced when the nucleus has excess of protons compared with the number of neutrons.

The general expression of the reaction is: $^{A}_{Z}X \rightarrow {}^{A}_{Z-1}Y + {}^{0}_{+1}e + {}^{+0}_{0}y$ (8),
it is produced one nucleus with an atomic number smaller than 1 and with the same number of atoms. In case of **β^{+} disintegration or in the presence of β**, two particles are formed with random energetic distribution. In this situation the emission spectrum is of continuous type.

When β^{+} is in contact with the electrons from an environment, it is produced a reaction **of annihilation**, accompanied with the release of two quanta of radiation **γ**; it is a feature of the β^{+} radioactivity.

The radioactivity γ. Contrary to the different types of radioactivity, which were mentioned, the radioactivity **γ** does not present an emission of mater from nucleus, but an emission of energy in form of electromagnetic radiation. It can be pure in the case of transitional passage from one metastable level to a stable one.

In the most of cases, the radioactivity **γ** (*fig. 17.6*) results from the unexcitation of the nucleus, caused by **α** or **β** disintegration.

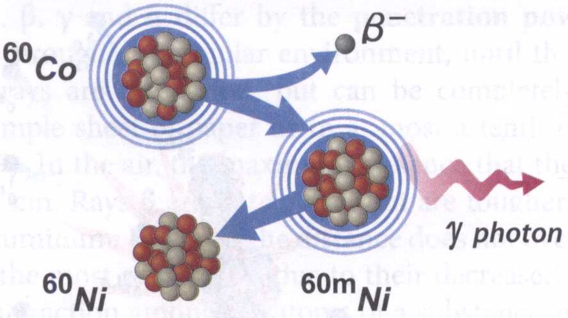


Fig. 17.6.

γ photons with an energy not more than a few MeV result from the transitions of nucleons among energetic levels that are perfectly defined in nucleus. Their spectrum lines originate from the nucleus. In some cases, the energy is given directly to one electron that is expelled. This phenomenon is of internal conversion. It determines the electronic rearrangement which spectrum is of lines.

The flow of neutrons consists an ionizing radiation and because of the interactions with atomic nuclei there are formed charged particles and photons γ. As example can serve the nuclear reactions caused by the capture of the neutrons:

$$ {}^{27}_{13}Al + {}^{1}_{0}n = {}^{24}_{11}Na + {}^{4}_{2}\alpha; \quad {}^{14}_{7}N + {}^{1}_{0}n = {}^{14}_{6}C + {}^{1}_{1}p. $$

The nuclear reactor (*fig. 17.7*), in which a chain reaction occurs, is the base of neutron production (*fig. 17.8*).

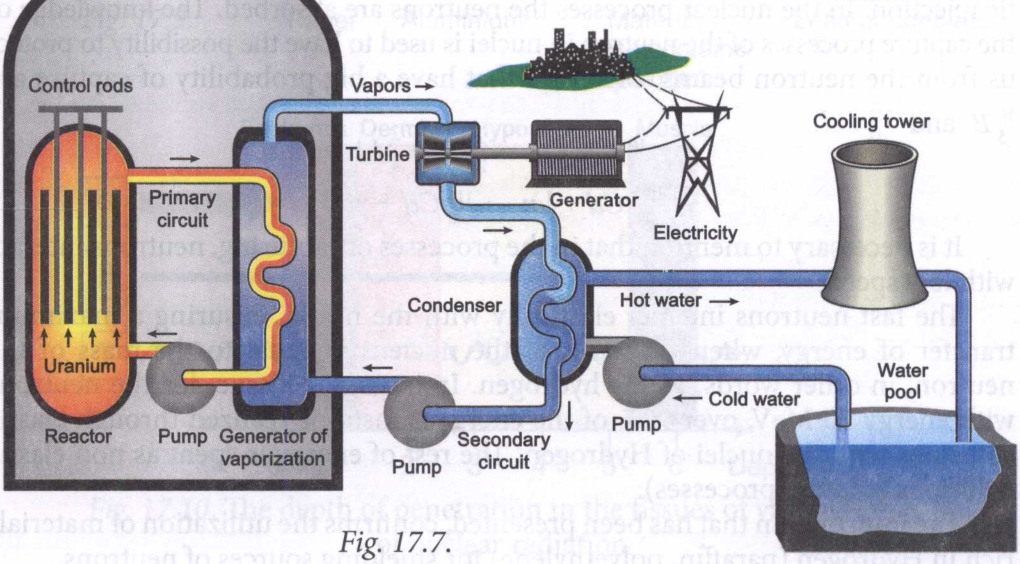


Fig. 17.7.

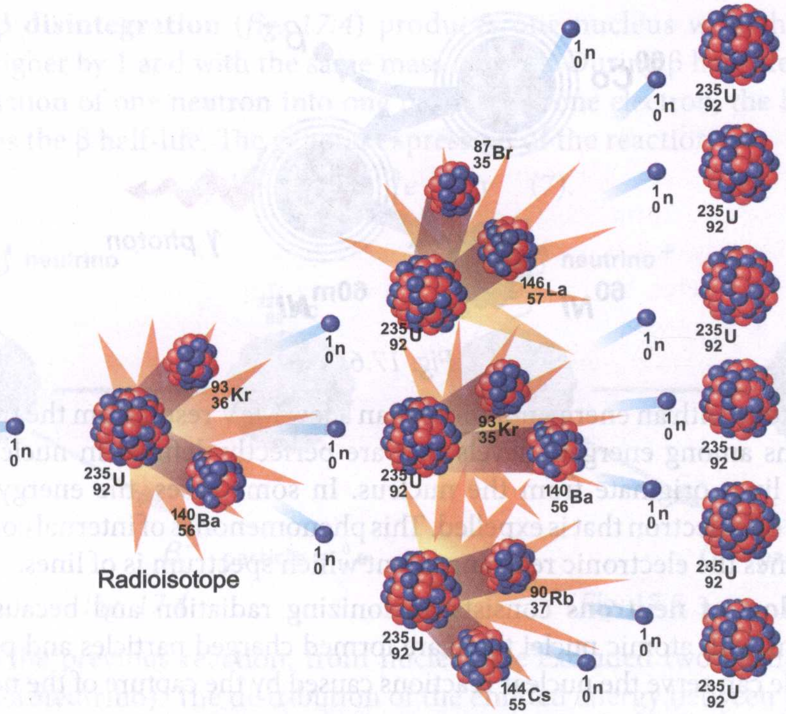


Fig. 17.8. Chain reaction.

The interaction between neutrons and the substance differs from the interaction between other particles, due to the fact that the neutron interacts with the nuclei, causing nuclear reactions. The probability of these reactions is very high, because neutron has no charge and its entrance in the nucleus is not held back by the electrostatic rejection. In the nuclear processes the neutrons are absorbed. The knowledge of the capture processes of the neutron in nuclei is used to have the possibility to protect us from the neutron beams. Materials that have a big probability of capture are $^{10}_{5}B$ and $^{113}_{48}Cd$:

$$^{10}_{5}B + ^{1}_{0}n \rightarrow ^{4}_{2}He + ^{7}_{3}Li$$

$$^{113}_{48}Cd + ^{1}_{0}n \rightarrow ^{114}_{48}Cd + \gamma .$$

It is necessary to mention that in the processes of capturing, neutrons interact with low speed.

The fast neutrons interact elastically with the nuclei, ensuring a maximum transfer of energy, when the mass of the nucleus is equal to the mass of the neutron, in other words, to the hydrogen. In biological tissue, for the neutrons with energy 10 MeV, over 90% of the energy transfer is realized through elastic collisions with the nuclei of Hydrogen. The rest of energy is spent as non elastic collisions (capture processes).

The information that has been presented, confirms the utilization of materials rich in Hydrogen (paraffin, polyethylene) for shielding sources of neutrons.

"Radiations" **α**, **β**, **γ** and **n** differ by the **penetration power** (the distance that they can pass through a particular environment, until they are completely absorbed). The *α* rays are the fastest, but can be completely stopped in the thickness of one simple sheet of paper or in at most a tenth of a millimeter of aluminum (*fig. 17.9*). In the air, the maximum distance that these rays can cross does not exceed 11 cm. Rays **β** are not so fast, but are tougher than **α**; they can cross 2–3 mm of Aluminum. In the air the distance does not overcome 10–15 cm. Instead, rays **γ** are the most penetrated, due to their decreased ionization power (the processes of interaction among the atoms of a substance are very rare). Not having electric charge and mass, rays **γ** can easily cross large thicknesses from animal and vegetable tissues, easy substances and a few centimeters from heavy substances as lead. Due to the big penetration power, bigger than has rays **X**, rays **γ** primarily are used with the same purpose. Thus, they are used in medicine and to measure some metal objects from exterior, without necessity to unfold them, to have control on some metal parts of manufacturing machines in series, in geological researches of prospecting the deposits of Petroleum and many others.

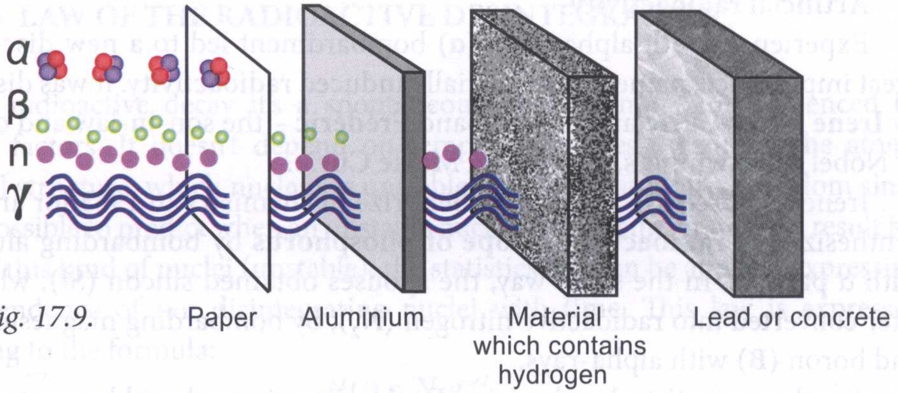


Fig. 17.9.

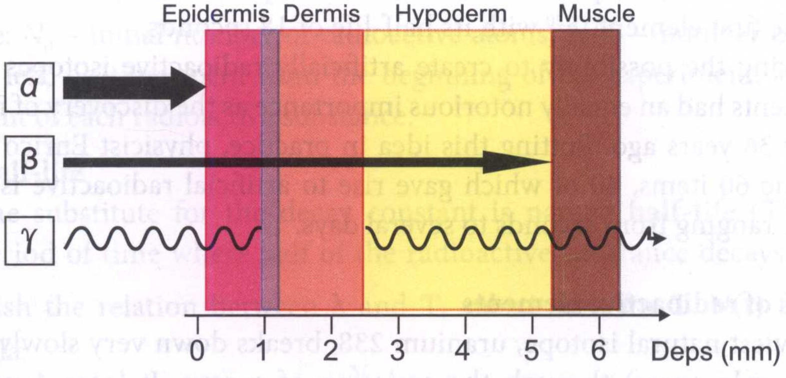


Fig. 17.10. The depth of penetration in the tissues of various types of nuclear radiation.

17.4. THE HISTORY OF RADIOACTIVITY

Natural radioactivity

The property exhibited by certain type of matter of emitting radiation spontaneously is called *natural radioactivity*. The phenomenon of radioactivity was first discovered in 1896 by the French physician **Henri Becquerel**, due to the general development of physics and due to the detection of X-rays in 1895 by **Röntgen**. In a lengthy process of modern scientific research, Becquerel observed that uranium emits invisible rays with similar properties to those of X-rays. The radiations similar to the X-rays are called *gamma radiations* or **gamma rays (γ)**.

Later, in 1898 **Marie Curie** and her husband discovered two other naturally occurring, strongly radioactive elements, **radium (Ra)** and **polonium (Po)**. At that time, there were also other erudits from all over the world working on this type of research.

Artificial radioactivity

Experiences with alpha- rays **(α)** bombardment led to a new discovery of great importance, named the artificially induced radioactivity. It was discovered by **Irene Joliet-Curie** and her husband **Frédéric** – the son-in-law and daughter of Nobel Prize winners, **Pierre** and **Marie Curie**.

Irene and Fréderic won the Nobel Prize in chemistry in 1935 for artificially synthesizing a **radioactive isotope of phosphorus** by bombarding aluminum with α particle. In the same way, the spouses obtained silicon **(Si)**, which was later converted into radioactive nitrogen **(N_2)**, by bombarding magnesium **(Mg)** and boron **(B)** with alpha-rays.

Another scientist, the physician **Enrico Fermi**, produced by neutron bombardment the first element ${}_9F^{20}$ with its half-life of 18 seconds.

Discovering the possibility to create artificially radioactive isotopes of the various elements had an equally notorious importance as the discovery of natural radioactivity 36 years ago. Putting this idea in practice, physicist Enrico Fermi bombed some 60 items, 40 of which gave rise to artificial radioactive isotopes with half-life ranging from seconds to several days.

Families of radioactive elements

The heaviest natural isotope, uranium 238, breaks down very slowly (half-life lasts 4,5 mln years) through the emission of **α-rays**. It loses 4 units of its atomic number, converting into thorium 234 – a radioactive element. Its nucleus emits a **β particle** and becomes the nucleus of ^{234}Pa (protactinium), which in turn converts into the nucleus of **^{234}U**, and produces a new isotope of uranium, named ***uranium II***. Uranium II transmutes into ^{230}**Th** – a brand-new isotope, with a special name, ionium. Then ionium desintegrates through alpha-emissions forming **radium ^{226}Ra**, which was discovered by the Curie. The decomposition of radium creates the **radon ^{222}Rn** – a heavy radioactive gaseous element. This sequence of radioactive decay processes goes on with the **polonium ^{210}Po**, also discovered by the Curie. At last, polonium converts into **the isotope of led ^{206}Pb** and the decay sequence ends with stable atoms.

The aforementioned radioactive elements form the ***radium-uranium family***. One of the radioactive family starts with the so-called *actini-uranium* ^{235}**U**, ending with a led isotope ^{207}**Pb**. The thorium family begins with ^{232}**Th** and ends with ^{208}**Pb**.

17.5. LAW OF THE RADIOACTIVE DESINTEGRATION

The radioactive decay it's a spontaneous phenomenon, not influenced by external factors. It doesn't depend on temperature, pressure or on the atom's chemical structure, which nucleus is unstable. This process is also a random since it is impossible to predict when an unstable nucleus will disintegrate. As a result for many of this kind of nuclei (unstable), the statistics law can be applied, expressing the dependence of not disintegrating nuclei with time. This law is expressed according to the formula:

$$N(t) = N_0 e^{-\lambda t}.$$

Where: N_0 – initial number of radioactive atoms; $N(t)$ – number of radioactive atoms left; t – time spent from the beginning of the experiment; λ – the decay constant of each radioactive substance.

Half-life

The substitute for the decay constant is named **half-life (T)**, represents the period of time where half of the radioactive substance decays. In order to establish the relation between **λ** and **T**, substitute **t** for **T**, **N(t)** for $\frac{N_0}{2}$ in the formula:

$$\frac{N_0}{2} = N_0 \cdot e^{-\lambda T}.$$

Simplifying the N_0 and logarithm, obtain: $T=\frac{\ln 2}{\lambda}$.

Considering the relation between λ and T, the decay law can be expressed by the following formula:

$$N(t)=N_0\cdot e^{-\frac{\ln 2}{T}\cdot t}=N_0\cdot(e^{-\ln 2})^{-\frac{t}{T}}=N_0 2^{-\frac{t}{T}}.$$

It becomes interesting when the time (t) contains an integer number of period ($t=nT$), where n is a positive integer. The mathematical expression obtained is:

$$N(nT)=\frac{N_0}{2^n}.$$

The half-life for different elements varies within very large limits (depending on the type of the element), – from a fraction of a second up to billions of years. Uranium-238, uranium-235, thorium-232 and potassium-40 have the longest half-life.

The law of radioactive decay is shown in figure 17.11.

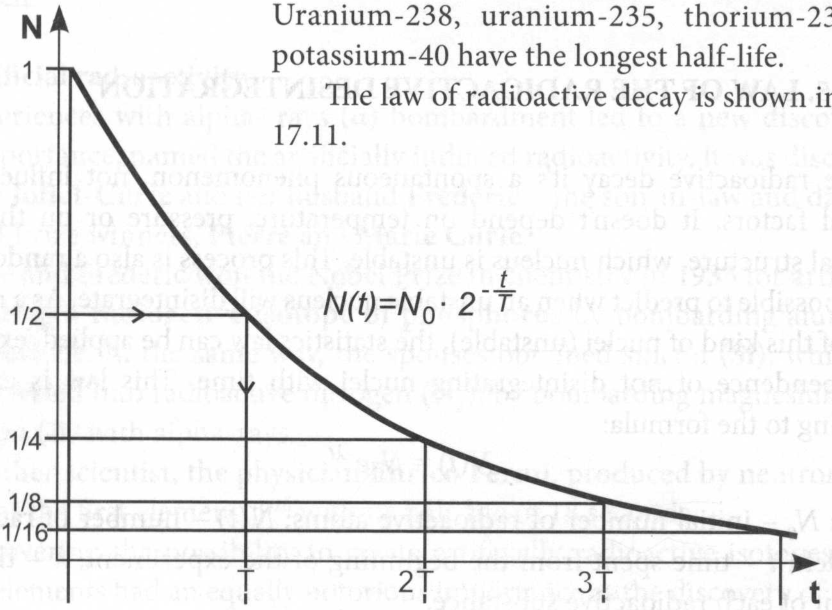


Fig. 17.11.

Radioactive activity

In order to work with radioactive sources, one has to know how many decays occur in a specific period of time. The number of radioactive decays represent an important particularity, called **radioactive activity (RA)**.

The unit of measurement of activity of a radioactive source in the International System of Units is **becquerels (Bq)**, which corresponds to activity of a radioactive source, in which it produce a disintegration into a second. In practice the most common unit is the **curie (abbreviated Ci)**:

$$1\ \mathrm{Ci} = 3.7\cdot 10^{10}\ \mathrm{Bq}.$$

The specific activity expressed in **(Bq/kg or Ci/kg)** represents the concentration of a radioactive substance in a particle.

Experimentally, the activity of a radioactive particle was determined by recording the number of particles or γ photons emitted from it during 1 second. Therefore, applying the notion of radioactivity, the decay law has mathematical expressions similar to the aforementioned:

$$A(t) = A_0 \cdot e^{-\lambda t};\ A(t) = A_0 \cdot 2^{-\frac{t}{T}};\ \text{and}\ A(nT) = \frac{A_0}{2^n}.$$

EXERCISES

1. Calculate the binding energy per nucleon of the following isotopes: ^{235}**U**. Calculate the energy, in joules that 1 gramme of uranium will release.

Answer: 7.4 MeV.

2. The desintegration of ^{235}**U** nucleus releases 163 MeV. Calculate the released energy of one gramme of uranium.

Answer: $67\cdot 10^9$ J.

3. Calculate in J the released energy at the preparation of 1 gramme of helium in the fusion reaction:

$$^2_1H + ^2_1H \rightarrow ^4_2He.$$

It is given: the atomic mass of deuterium = 2.014102 u.m.a.;
the atomic mass of helium = 4.002603 u.m.a.

Answer: $5.75\cdot 10^{11}$ J.

4. Calculate in Joule the released energy from the annihilation of a positron into an electron.

Answer: $1.64\cdot 10^{-13}$ J.

5. The activity of ^{131}**I** and ^{60}**Co** are equal to 1 mCi? Calculate the mass of each isotope.

It is given: $N_A = 6.023\cdot 10^{23}\ \mathrm{mol}\cdot l^{-1}$;
period ^{131}I = 8.08 days;
period ^{60}Co = 5.24 years.

Answer: $m_I = 8.1\cdot 10^{-9}$ g; $m_{Co} = 8.7\cdot 10^{-7}$ g.

6. How many days does the activity of a source ^{131}**I** take to decrease from 200 kBq to 25 kBq? $T_I = 8$ days.

Answer: after 24 days.

7. Initially, the activity of an isotope is 5 kBq. After 10 days, the isotope's activity decreases to 3.08 kBq.

Calculate the half-life (time taken for half the radionuclide's atom decay).

Answer: 14.3 days.

8. A bottle of iodine 131 has 100 mCi at moment t = 0. Calculate it's activity after a period of 48 days. It is known that $T \approx 8$ days.

Answer: 1.56 mCi.

9. Calculate the radioactivity (Bq) of a source that contains 1g of $BaCl_2$ with radioactive $^{133}_{56}Ba$ (T = 8 years) and stabile $^{36}_{17}Cl$?

It is given: $N_A = 6.023 \cdot 10^{23}\ mol \cdot l^{-1}$.

Answer: $8.11 \cdot 10^{12}$ Bq.

10. The radioactivity of 1 mg of **NaI** equals to 3920 MBq. NaI contains stabile $^{23}_{11}Na$ and radioactive $^{131}_{53}I$. Calculate the period of iodine 131, knowing that $N_A = 6.023 \cdot 1023\ mol/l$.

Answer: 8 days.

11. After the bombardment with alpha rays, the nucleus of ^{210}Po converts into nucleus of ^{206}Pb. Determine the energy of α desintegration.

Answer: 5.40 MeV.

12. Radium ($^{226}_{88}Ra$) is a strong radioactive element. When radium decays, it creates radon (Rn). The atomic mass of radium is 226.0254 u.a.m. Determine the mass deviation and the binding energy per nucleon for radium.

Answer: $\Delta m = 1.8176$ u.m.a.; B = 7.49 MeV.

13. The radioactive elements, A and B, have the following decay constants: λ_A and λ_B. Determine the relation between T_A and T_B, knowing that $\lambda_A = 2\lambda_B$.

Answer: $T_B = 2T_A$.

14. The unstable isotope of ^{10}C (Z = 6) when disintegrates becomes Bohr (Z = 5). Initially, at t = 0, the radioactive activity of a substance is $1.6 \cdot 10^4$ Bq. One minute later, this activity is equal to $2 \cdot 10^3$ Bq. Determine, for carbon C, the half life.

Answer: T = 20 s.

15. The radiaoctive element $^{99m}T_C$ with half-life of (T = 6 hours) is applied in medical diagnosis. An examination requires injecting a solution of 260 MBq. Determine the activity of the substance in pacient's organism after two days.

Answer: A = 1.02 MBq.

16. One ml of seawater contains 10^{-15} g of radon and an activity of 10 mCi. Determine the quantity of water (m^3), if the activity of 1 g of radon equals to $6.46 \cdot 10^{-6}$ Ci.

Answer: $15 \cdot 10^{11}$ m^3.

18. DETECTION DEVICES. FUNDAMENTALS OF DOSIMETRY

18.1. OPERATING PRINCIPLES OF NUCLEAR RADIATION DETECTORS

Ionizing radiation used for different purposes is highlighted or measured using **detectors** coupled with auxiliary systems, which allow recording and presenting them in the form of curves or pulse. When necessary, the detectors can identify some characteristics of radiation, such as the type of radioactivity, energy of photons or particles, particle mass.

Depending on the operating principle are known several types of detectors, but the most used are **the detectors that are based on ionization phenomena in gases.** Such detector is **ionization chamber** (*fig. 18.1*).

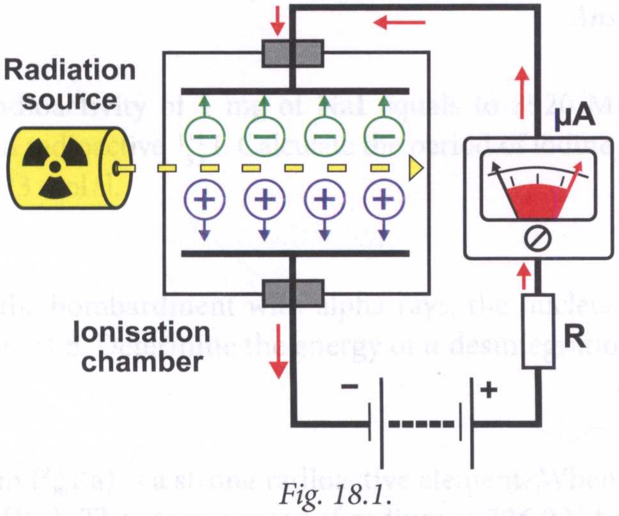


Fig. 18.1.

Into a closed chamber two electrodes are mounted isolated from the body of the chamber and connected to a continuous current source. The space of the camber is filled with gas, more often with air at normal pressure. The radiation source (with low penetration power, for example **α-radiation**) is situated inside the chamber or (with a high penetration power, for example **γ-radiation**) outside of it. Whens passing through the chamber, the particles produce pair of ion-electron that are collected by electrodes. The potential difference between the electrodes should be sufficiently high, to able the microelement from circuit to record the saturation current. Only in this case is excluded the process of recombination of ions and the ionization current is proportional to the total number of ion-electron pairs, created by particles in a unit of time, and, therefore with the same activity of the radiation source. In an ionizating chamber the current is very low, which leads to complicated registration system. Therefore, the most frequent used another detector – **Geiger–Müller counter**, containing gas (*fig. 18.2*).

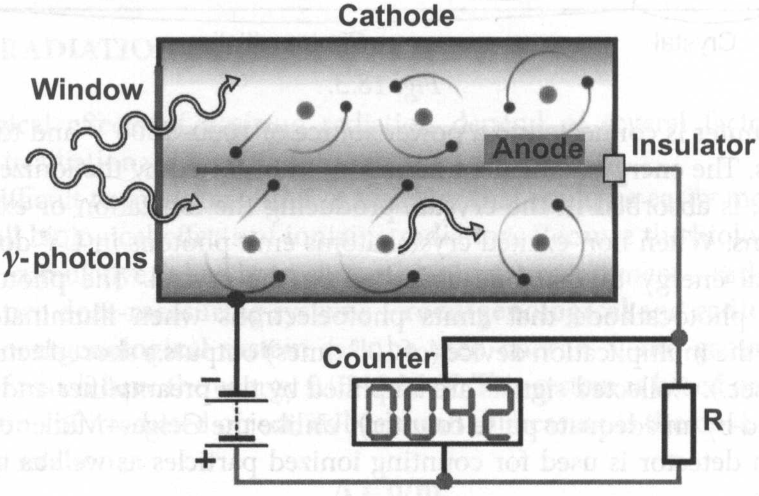


Fig. 18.2.

Geiger-Muller counter is composed of a condenser enclosed in a cylindrical glass tube that contains a mixture of argon and alcohol vapors at low pressure. The **anode** is a very thin thread (0,1 mm), the **cathode** is a metal drum or a metal film deposited in the inner wall of the glass tube. The electrodes are connected to a potential difference of about few hundred of volts. When the particle passes through the gas (inside the counter), an ion-electron pair is formed. Electrons are strongly accelerated in the intense field of anode and produce **secondary ionization**. In this way is formed an avalanche of electrons and positive ions. In circuit occur a short current that produces at the resistance terminals (**R**) decrease in tension, which is amplified and recorded. This is **voltage pulse** obtained at the passing of ionized particles through the counter. Geiger–Müller counter records only the number of particles, but doesn't allow to determine their other characteristics.

Detectors that are based on the appearance of flicker

Scintillation detector consists of crystal a **NaI**, fluorescent, thallium (**Tl**) activated and coupled to a photomultiplier that supplies with electrical impulses the preamplifier (*fig. 18.3*).

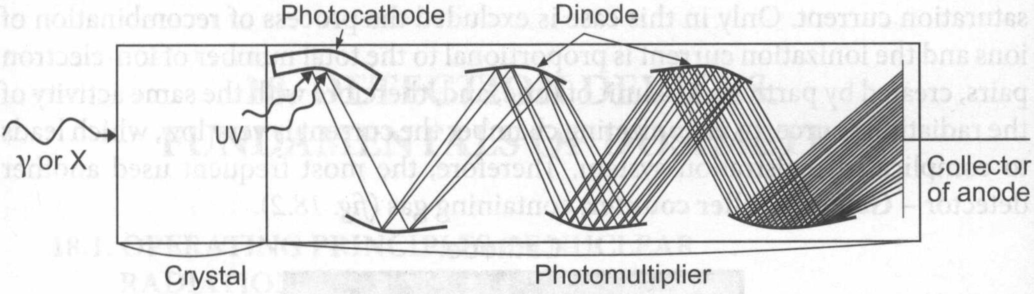


Fig. 18.3.

The counter is connected to a power source of 1000–1500 V and to a counter of impulses. The energy of incident radiation, transported by the ionized particles or photons, is absorbed in the crystal, producing the ionization or excitation of crystal atoms. When non-excited crystal atoms emit photons in **UV** domain with proportional energy to that one absorbed by the crystal. The photomultiplier contains a photocathode that emits photoelectrons when illuminated by the crystal, but the multiplication device (of 10^{10} times) outputs a short electric impulse (10^{-6}–10^{-8} sec.). Collected signals are amplified by the preamplifier and processed (are counted by an adequate pulse counter). Unlike the Geiger–Müller counter the scintillation detector is used for counting ionized particles as well as to measure their energy.

Detectors that are based on the phenomenon of forming pairs electron–gap in semiconductor crystals

Through their structure, the semiconductors contain a certain number of free electrons, which can move inside the substance. If through such a semiconductor crystal, situated between two metal electrodes (*fig. 18.4*), ionizing radiation passes, crystal ionization occurs, and due to electron- gap pairs, considerable increases the electrical current through crystal. Semiconductor crystals from detectors are

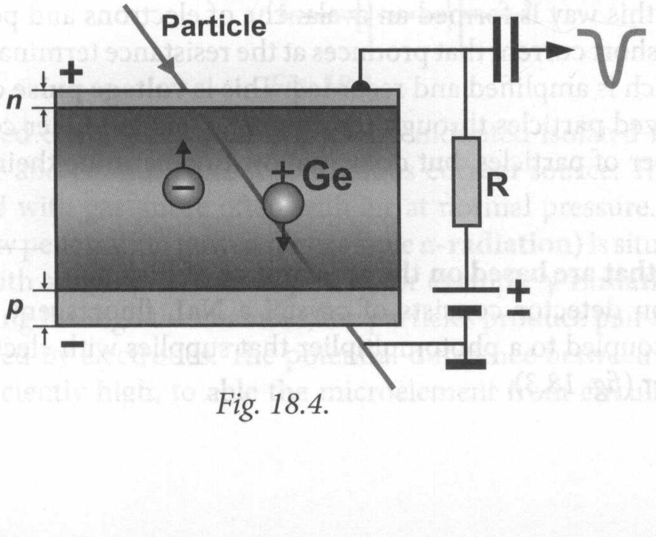


Fig. 18.4.

made of **cadmium sulfide.** The number of electron–gap pairs is proportional with the ionizing particle energy. The electric impulse amplitude is proportional to the energy of recorded nuclear particles. The main qualities of semiconductor crystal detector are: small size, very good energy resolution and the fact that can be used for all types of nuclear radiation (α, β, γ).

18.2. RADIATION DOSE

Biological effects of ionizing radiation depend of several factors: type of radiation, irradiation geometry, etc.

It is difficult to mention a unit of measure that could be easily measured and valid for all biological effects of ionizing radiation. Because the biological effects are largely cumulative, it had been chosen a unit of measurement – **radiation dose.** There are two dose-measuring systems – **roentgenological** and **radiobiological**.

The roentgenological system may be used only in X and gama radiations, photons of which have the energy 0.03–3 MeV. This system is based on measuring the ionization of irradiated object; its basic unit of measure is **the dose of ions (Δ)**, defined by relation:

$$\mathbf{\Delta = q/m.}$$

Where: **q** – total electric charge of the ions of both signs; **m** – object mass.

The unit of measurement for dose of ions in S.I. is coulomb/kilogram ($C \cdot kg^{-1}$). But, for practical reasons, it is used the unit called – **roentgen (R).**

By definition, roentgen is the dose of X or γ radiation, that in one kg of air, under normal conditions produces a number of ions, with total charge of $2.58 \cdot 10^{-4}$C.

Therefore, **$1R = 2.58 \cdot 10^{-4}\,C\,kg^{-1}$.**

The dose rate (δ) is the ratio between radiation dose and the time of irradiation of the object:

$$\delta = \frac{\Delta}{t} \text{ (R/s; R/min; R/h).}$$

The radiobiological system is based on determination of the energy absorbed by the irradiated object. The basic unit of measure of this system is the **absorbed dose (D)**, defined as absorbed energy by a unit of mass:

$$D = \frac{W}{m}.$$

Where: **m** – mass of substance or irradiated tissue; **W** – absorbed energy in this mass.

The unit of measurement in S.I. of dose of absorption is **Gray (1 Gy = 1 J/kg)**.

It was found that the same energy absorbed by ionization radiation of different nature produces different biological effects. This made necessary to introduce in the radiobiological system a coefficient called **relative biological effectiveness (Q)** (quality factor). The product of the absorbed dose (**D**) and the quality factor (**Q**) represents **biological dose** (**B**):

$$\mathbf{B = Q \cdot D}.$$

Q = 1 for β and γ radiation;

Q = 3–10 for neutrons (depends on their energy);

Q = 10 for α particles.

Unit of measurement for biological dose in S.I. is called **Sievert (Sv)**.

Another often used unit of measure is **biological dose rate (b)** given by relation:

$$\mathbf{b = B/t}.$$

Where **t** is time of irradiation.

Integral biological dose (B_i) is given below:

$$\mathbf{B_i = B \cdot m}.$$

Where **m** is mass of the irradiated object or tissue.

The human body is permanently or occasionally irradiated by couple sources of ionizing radiation: natural, sanitary and technical etc.

Natural irradiation has natural background (environmental fund), composed of 30% cosmic radiation, 45% comes from the activity of radioelements from the soil, 24% of radiation comes from metabolism and 1% from radioactivity of atmosphere.

Fundamental safety rules of radiobiology, drafted by the **National Commission of Romania** for Control of Nuclear Activity (29 august 2000), provide certain values of maximum annual allowed dose (*table* 2).

Table 2

Irradiated object	The maximum dose permitted	
	Professional exposure	**Population**
The whole body	20 mSv/year	2 mSv/year
Crystalline	150 mSv/year	15 mSv/year
Skin	500 mSv/year	50 mSv/year
The extremities limbs	500 mSv/year	50 mSv/year

18.3. BIOLOGICAL EFFECTS OF RADIATION AND PROTECTIVE MEASURES FOR RADIATION

As a result of interaction between radiation and living organism occur physical phenomena (ionization, excitation) that determine chemical phenomena (alternations of macromolecules and of enzymatic systems). But the most important effects can be observed in germ cells. After interaction between radiation and germ cells can be observed a disruption of chromosomes and genetic code – DNA. The severity of this problem is amplified by transmission to descendants even in very low doses.

For those who work in radioactive environment one of the most simple solution to decrease the absorbed dose is to wrap with protector screen (from lead) the devices that use radiation. The activity with any kind of sources of ionizing radiation needs to protect the personal against its damaging action. Conventionally are three types of protection: through time, through distance and through shielding materials. In figure 18.5 are presented the informational warning symbols for radioactive substances.


a) b)

Fig. 18.5: a) new hazard warning symbol for highly radioactive elements; b) hazard warning symbol of radioactive substances stored in protective containers.

The flow of ionizing radiation can be reduced by interposing screen protector between the source and the irradiated object.

This method is based on the law of attenuation of the radiation intensity passing through a layer of substance.

The effectiveness of the mitigation screen (E) can be determined from the relationship:

$$E = \frac{A_0 - A}{A_0}.$$

Where: A_0 – the initial activity of the source in the absence of screen; A – source's activity with the screen.

EXERCISES

1. A point source of gamma radiation has a dose rate of 200 Gy/hour at the distance of 1 meter. Calculate the dose rate at a distance of 2 meters from the source.

Answer: 50 Gy/hour.

2. During 6 hours, a body of mass 60 kg absorbed $W = 1$ J energy of ionizing radiation. Determine the absorbed dose in Gray.

Answer: 0.017 Gy.

3. In a city, the radioactive fund is 30 mR/h. Determine the ions dose received by each citizen of this city during one year.

Answer: $0.676\ \mathrm{C \cdot kg^{-1}}$.

4. In 10 g of water were absorbed 10 protons with 5 MeV energy. Determine the absorbed dose.

Answer: $8 \cdot 10^{9}$ Gy.

5. The α particle produces in the air $15 \cdot 10^{4}$ pairs of ions. Determine the ionization current produced by a source that emits 100 α particles per second.

Answer: $2.4 \cdot 10^{-12}$ A.

6. Passing through a substance layer with thickness of 5 cm, the gamma radiation intensity had decreased 6 times. Determine the linear attenuation coefficient of this substance.

Answer: $0.36\ \mathrm{cm^{-1}}$.

7. The ion dose rate at the distance of 10 cm from the source is 85 mR/h. At what distance from the source we can stay, without protection, if the allowed dose is 0.017 mR/h?

Answer: 70.71 m.

8. The activity of a radioactive source ^{60}Co, used in radiotherapy, is 10^{12} Bq. The produced dose rate of ^{60}Co at the distance of 1m is $3.5 \cdot 10^{-13}\ \mathrm{Sv \cdot hour^{-1}}$. The individual dosimeter of the operator, whose place of work is at 5 m from the source, shows 1 mSv. Determine the time of exposure.

Answer: 4.35 min.

19. PARTICLE ACCELERATORS. PHYSICAL BASIS OF RADIOTHERAPY

19.1. PARTICLE ACCELERATORS

General concepts

It is known that for the acceleration of a body, it must be applied a force.

In case of particle accelerators, force of acceleration has electrical nature, it cannot be applied rather than only on electrified particles: **e**$^-$, **p**$^+$, **α**$^{++}$ or different ions.

Applied forces are of two types:

- electrical;
- magnetically.

Electrical forces

By definition, the force that the electric field applies on an electric charge, at a given point, corresponds to the intensity of the field ($\vec{E}$). If $\vec{E}$ is the same in any point of the field, it is said that the field is uniform.

The movement of a charged particle **+q** in a longitudinal electric field, created by a potential difference (**V**) is represented in figure 19.1.

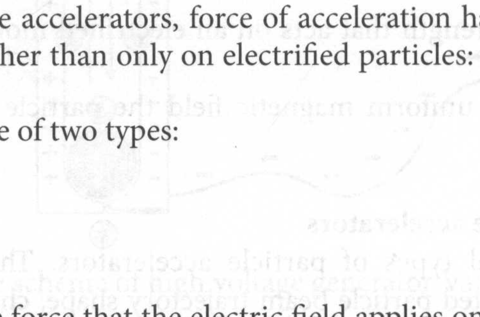


Fig. 19.1.

On the particle acts the force:

$$\vec{F} = qE; \quad E = \frac{V}{l}; \quad F = q\frac{V}{l}.$$

Under the action of this force, when crossing the potential difference (V), particle with charge q gains kinetic energy by the amount:

$$W = q\,V = \frac{1}{2}\,mv^2.$$

Where: v – final speed; **m** – particle mass.

In case of transversal electric field, initial speed remains constant in module, varying only in direction.

Deviation of motion of the particle in relation with the original direction is inversely proportional with initial speed and particle mass.

Magnetic forces

Magnetic field strength that acts on an electrified moving particle is called **Lorentz force.**

In an unlimited, uniform magnetic field the particle will present circular trajectory.

Different particle accelerators

There are several types of particle accelerators. They can be classified according to: accelerated particle beam trajectory shape, character of fields accelerators, printed particle energy range and the nature of accelerated particles. We will refer only to some types of accelerated particles used more often in medicine.

a. High Voltage electrostatic accelerators Van der Graaf

This type of accelerator needs a big potential difference. It is composed of two elements:

– a generator of high tension (*fig. 19.2*);

– a vacuum acceleration tube.

The scheme of electrostatic accelerator of direct action is represented in figura 19.3. The particles or nuclear ions are accelerating due to direct crossing of a large potential difference, 5–6 MV. In the accelerating tube is achieved maximum possible vacuum (10^{-5} – 10^{-6} mm Hg).

Particle with charge Ze acquires kinetic energy W = ZeV. The advantages of accelerators of direct action are: high tension, continuity, stability of accelerated energy flow, where current can reach a few milliamps. They can be used in nuclear physics and radiotherapy, with the help of accelerated electrons, can be obtained X-rays of high energy. But, in these accelerators the energy of particles is not above 50 MeV. In order to obtain a higher energy of particles, are used **linear accelerators**.

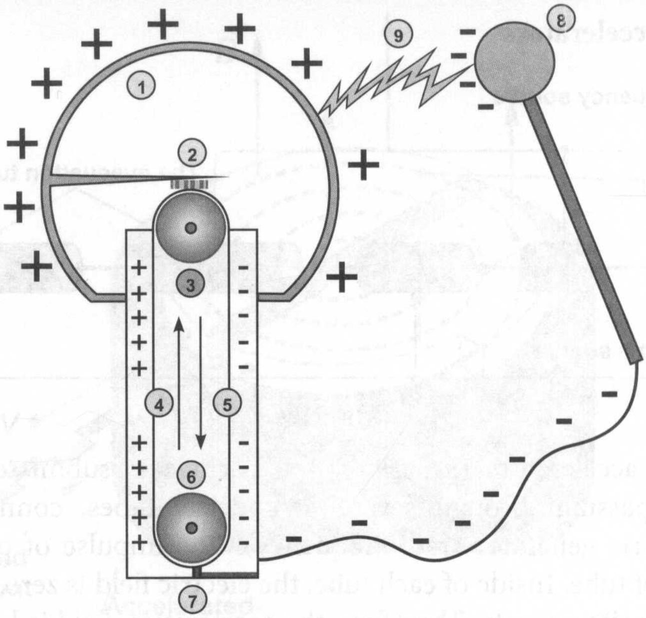


Fig.19.2. The scheme of high voltage generator Van der Graaf:
1) empty metal scope; 2) superior electrode; 3) superior cylinder (metallic); 4) the positively charged belt; 5) the opposite negatively charged belt; 6) inferior cylinder(acryl glass); 7) inferior electrode (soil); 8) the negatively charged sphere used to uncharged the main field; 9) spark produced by the difference of potential.

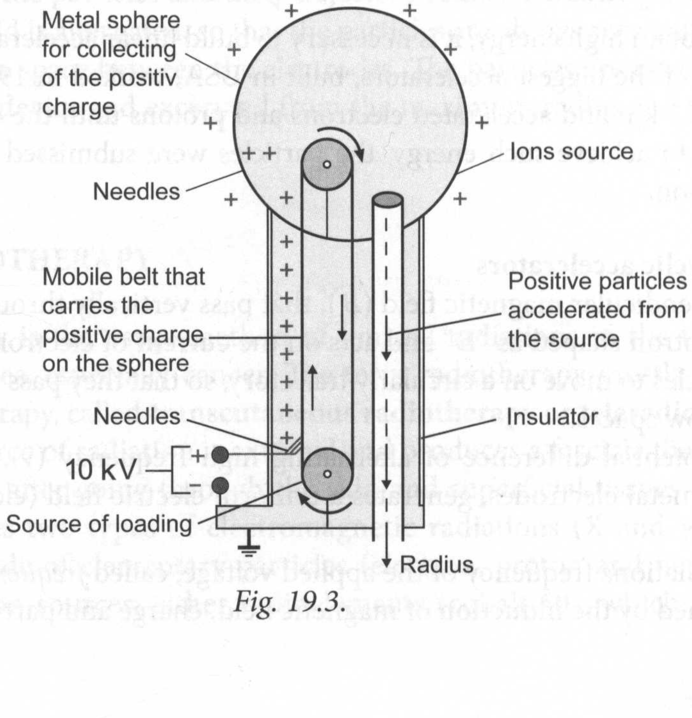


Fig. 19.3.

b. Linear accelerators

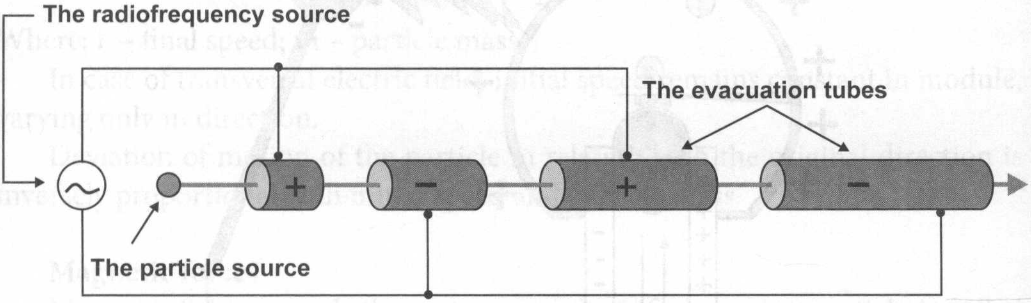


Fig. 19.4.

In linear accelerators (*fig. 19.4*), the particle is submissed to multiple accelerations, passing through several cylindrical tubes, connected to high frequency electric generator (radiofrequency). The impulse of particles moves along the axis of tube. Inside of each tube, the electric field is zero. Neighbouring tubes have opposite polarity. Therefore, the acceleration field is located between the tubes. The frequency of the generator and the size of the tubes must correspond to the conditions, so that the pulse of particles occurs in the next period when the tubes polarity invert. For this, the length of the tube (l), the speed of the particle (V), and the frequency (ν) of the field must correspond to the relationship:

$$l = \frac{V}{\nu}.$$

To obtain high energy, it is necessary to build linear accelerators of big length.

One of the biggest accelerators, built in USA, worked in 1989–1998. It had a length of 3 km and accelerated electrons and protons until the energy of 50 GeV. In order to achieve such energy, the particles were submissed to 80 000 acts of acceleration.

c. Cyclic accelerators

Perpendicular magnetic field ($\vec{B}$), that pass vertically through the electrodes of a cyclotron shaped as “D” and acts on the current of electrons or ions, forcing the particles to move on a circularly trajectory, so that they pass repeatdly through the narrow spaces.

A potential difference of alternating high frequency (ν), applied between the two metal electrodes, generates a uniform electric field (electric field is zero inside).

Oscillations frequency of the applied voltage, called *frequency of cyclotron*, is determined by the induction of magnetic field, charge and particle mass:

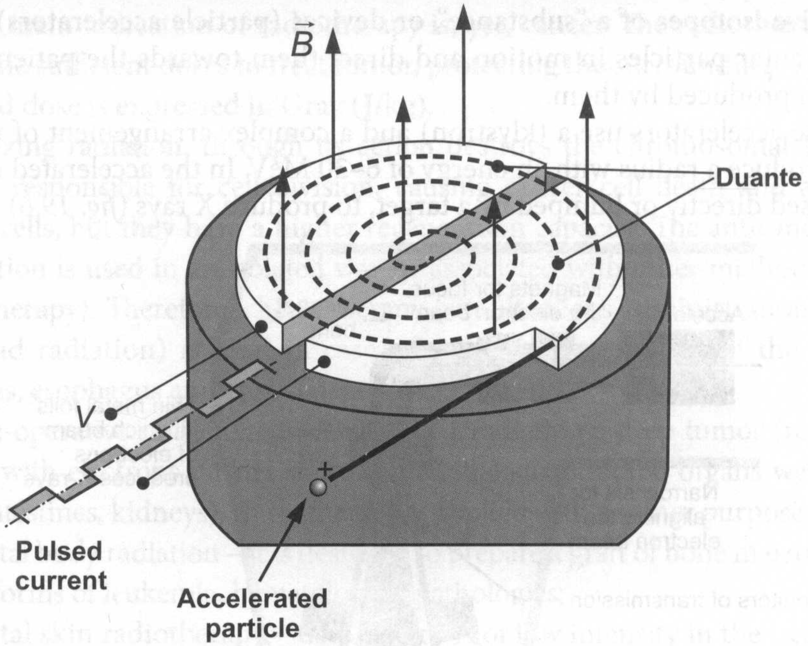


Fig. 19.5. Simplified diagram of cyclotron.

$$\nu = \frac{Bq}{2\pi m}.$$

The speed of the particles and their trajectory radius increases gradually the polarity of the field is alternated, so that the particles are always accelerated when they pass through space between the electrodes. The particles are introduced in the center of the device and excerpted from the maximum radius and speed (or energy).

19.2. RADIOTHERAPY

Radiotherapy is the usage method of ionized radiations in the treatment of certain illnesses, especially cancer. The term radiotherapy mostly refers to external radiotherapy, called **transcutaneous radiotherapy** or **teleradiotherapy**, in which the source of radiation is external and produces a fascicle that reaches profound tissues after going through the skin and superficial tissues. External radiotherapy uses two types of electromagnetic radiations (X and γ) as well as radiations mode of elementary particles (electron, proton and neutron). It uses two radiation sources: either radioelements (cobalt 60), which are often radioactive isotopes of a "substance", or devices (particle accelerators) that set the elementar particles in motion and direct them towards the patient, or the radiation produced by them.

These accelerators use a (klystron) and a complex arrangement of magnets, which produce a radius with an energy of 6–30 MeV. In the accelerated electrons can be used directly or bumped by a target, to produce X rays (*fig. 19.6*).

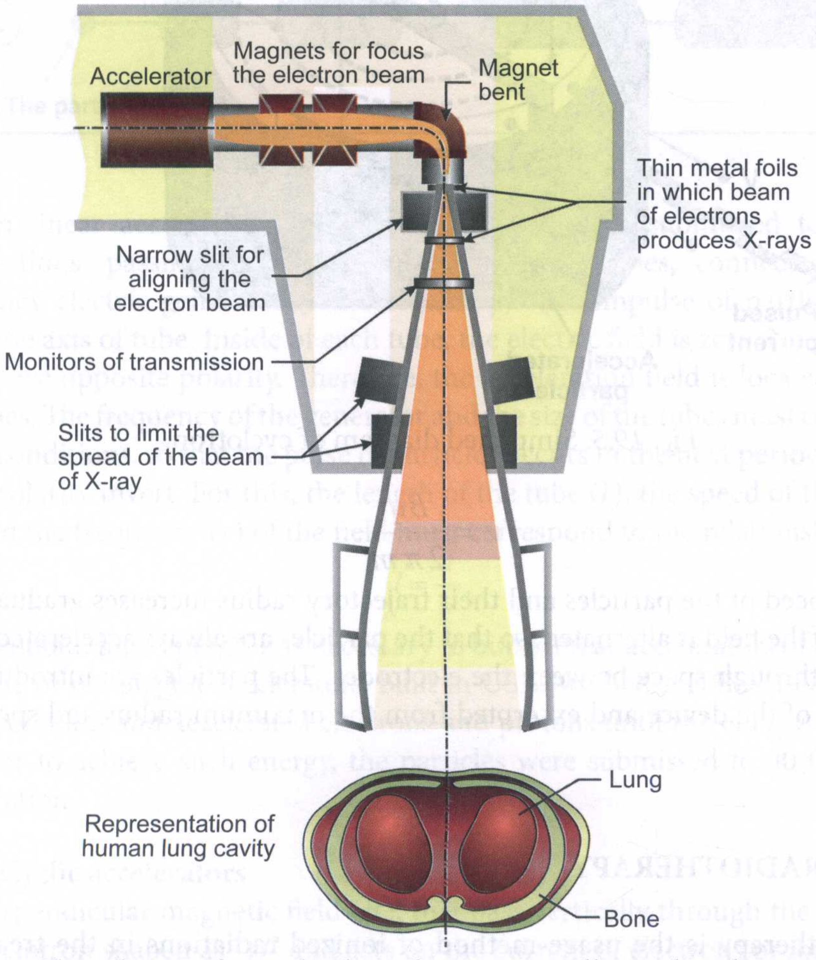


Fig. 19.6. The scheme of a medical accelerator typically used in therapy of cancer.

Safety, flexibility and accuracy of the product replace the cobalt-60 therapy as a treatment instrument.

Indications

At low doses, radiotherapy has an inflammatory effect, used sometimes in the treatment of shingles or kelloids (pathological scars).

The main indication of radiotherapy is, yet, cancer. The radiotherapy aim is to provide sufficient doses to treat tumor, protecting the surrounding organs. The absorbed dose is expressed in Gray (J/kg).

Ionizing radiation, through its action destroys the chromosomal structures that are responsible for cell division, causing cancer cell death and attack the healthy cells, but they have a higher regeneration capacity. The anticancer action of radiation is used in an isolated way or associated with other method (surgery, chemotherapy). Therefore, chemotherapy (simultaneous administration of medicines and radiation) is used in case of epidermal carcinoma of the pharynx, bronchus, esophagus and anal cancer.

Pre-operative radiotherapy consists of irradiating a deep tumor (rectal, pancreatic) with electrons during surgery, after the surrounding organs were moved aside (intestines, kidneys). Radiotherapy is applied with another purpose:

– total body radiation – it is designed to prepare a graft of bone marrow to treat certain forms of leukemia, hematological pathologies;

– total skin radiotherapy – uses electrons of low intensity in the treatment of cutaneous lymphomas.

Different types of devices

Traditional tubes that produce **X** radiation of low energy are not used anymore except in the treatment of skin cancer. Linear accelerators are the most used devices: they produce active surface electrons and indicated in the treatment of superficial tumors or **X** radiation of high energy, which gets deep under the skin and are indicated to treat of deep cancer of chest and abdomen. Another variety of particle accelerators, **cyclotrons**, that produce neutrons and protons, are used in the treatment of some rare forms of cancer that requires caution (eye melanoma, skull base sarcoma). And, in the end devices that contain cobalt emit gamma radiation, and treat the deep tissue, but don't affect the skin. They are applied in the treatment of head, neck, chest and limbs cancer.

Technology

Modern radiotherapy assumes an expensive technical environment. A prior examen to the scanner allows finding the location of the irradiated organs. It is performed a simulation by simple radiography to allow proper positioning of the beam. The necessary radiation dose for the destruction of cancerous cells is calculated through dosimetry. To increase the effectiveness of radiation without harming healthy tissues is used convergent beam technique: a bean irradiate the front side of the diseased part, second – the backside, third- the right side and last – the left side. Each beam has a very weak intensity to damage the healthy tissues in this path, but the sum of intensities of the four beams has a marked effect on tumor.

Irradiation through multiple convergent minibeams (*fig. 19.7*) is used in the treatment of arterivenous malformations or brain-limited tumors, but inoperable. The unwished effects of radiotherapy are the damage produced to healthy tissues. The main prevention of this effect is based on technical precautions: the radiation dose and the volume of the irradiated body to be as small as possible, to decrease the administrated dose per session, but to increase the number of sessions.

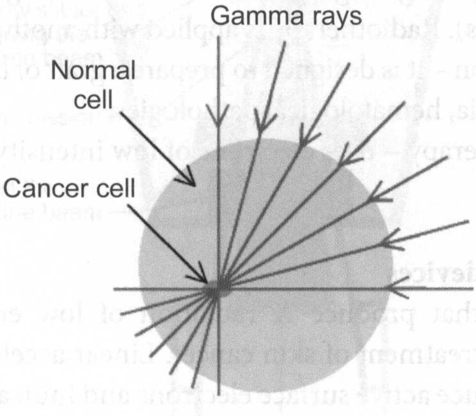


Fig. 19.7.

The most recent radiotherapy

Radiotherapy is based on photons, largely used nowadays, that allow the destruction, through a treatment session, of only 10% of irradiated tumor cells. Because of this, the specialists keep attention on the high risk of restoration of other cells.

A new type of radiotherapy, called **hadrontherapy**, which exists in Japan, USA, Russia and in other European countries, has the capacity to destruct almost all the irradiated cells.

Hadrontherapy consists in using hadrons (mostly proton and ion of carbon) in order to eliminate the cancerous tumors. One of the main advantages of this kind of therapy, in oncology domain, is its effectiveness in the case of tumors that cannot be treated using the conventional radio-therapeutically treatment.

Plus, the protons and ions of carbon have a higher precision than photons, and this makes hadrontherapy to destruct the malignant tissues, but sparing the healthy ones.

Another quality of hadrontherapy is that there is no risk of appearance of radio-resistance this means that the cancerous cells won't gain resistance to the treatment if using protons and ions of carbon. Hadrontherapy is applied in malignant inoperable tumors, mostly for those that are situated in the skull region, salivary glands, sinuses, prostate and eye region.

EXERCISES

1. What should be the intensity of electric field between the plates of a capacitor situated horizontal, to maintain in balance an electron?

Answer: $5.7 \cdot 10^{-11}\ \mathrm{Vm^{-1}}$.

2. Determine the kinetic energy of an electron that has a speed of $5 \cdot 10^{9}\ \mathrm{cm \cdot s^{-1}}$.

Answer: $7.1 \cdot 10^{-3}$ Mev.

3. Determine the acceleration voltage that allows obtaining electrons with a speed of $5 \cdot 10^{9}\ \mathrm{cm \cdot s^{-1}}$ in a Van der Graaf accelerator.

Answer: 7.1 kV.

4. Determine the speed of an accelerated electron by a potential difference of 4.55 kV in a Van der Graaf accelerator.

Answer: $4 \cdot 10^{7}\ \mathrm{m \cdot s^{-1}}$.

5. Calculate the force exerced by a uniform magnetic field with $B = 10^{-2}$ T on an electron that moves with speed of $2 \cdot 10^{7}\ \mathrm{m \cdot s^{-1}}$ perpendicularly to the surface of the field.

Answer: $8.8 \cdot 10^{-30}$ N.

6. What should be the tension applied between the duantes of a cyclotron, for the particles with the ratio $\frac{q}{m} = 9.6 \cdot 10^{7}\ \mathrm{C \cdot kg^{-1}}$ to move from the first round on a circularly trajectory with radius of 0.5 m?

Answer: $12 \cdot 10^{3}$ kV.

7. A linear accelerator works from a high-frequency alternating current $\nu = 5\cdot10^6$ Hz. The electrons which must be accelerated are injected at the entrance to the first electrode when t = 0, with initial velocity (v_0) of 10^6 m/s^{-1}. Determine the length of the first electrode.

Answer: 10 cm.

8. In a linear accelerator, an electrode with initial energy (E_0) continues to move between two cylindrical electrodes: one with zero potential, and the other with the potential $V>0$. Determine the velocity of the electron inside of the second electrode.

Answer: $v = \sqrt{\dfrac{(2E_0 + eV)}{m}}$.

9. For a course of radiotherapy were used β^- particles with energy of 4 MeV and linear movement on a tissue of 1 cm. knowing the energy expended in forming an ion pair, determine the transfer of linear energy (TLE) and the linear density of ionization (LDI).

Answer: TEL = $4\cdot10^6$ eVcm^{-1}; DLI = 10^5 cm^{-1}.

20. AUDITORY ANALYZER

20.1. GENERAL INFORMATION ABOUT SOUNDS

Sound wave

The sounds are longitudinal mechanical waves, in other words, an oscillating mechanical motion, which is transmitted through the environment. Unlike electromagnetic waves, sound waves cannot propagate in vacuum.

Sound waves require environmental particles to oscillate relative to the equilibrium position, parallel to the direction of wave propagation. In figure 20.1 is represented the analogy between longitudinal and transverse waves.

Sound sources are objects capable of transmitting mechanical vibrations corresponding to the environment surrounding them.

It is obvious that the source frequency matches to the frequency of the sound wave.

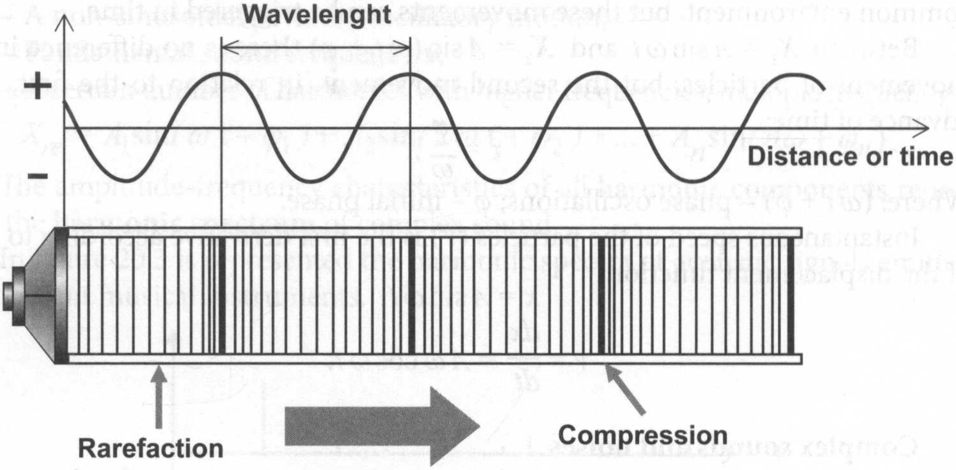


Fig. 20.1.

The sounds perceived by the human ear have frequency between 16 Hz and 20 KHz.

In the air, depending on the atmospheric pressure, the velocity of propagation of acoustic waves varies between 330 and 340 $m \cdot s^{-1}$.

The velocity of sound depends on the environment density – in water, for example, it is 1450 $m \cdot s^{-1}$; in bones – 3300 $m \cdot s^{-1}$.

Pure sounds

They are the simplest sounds. The longitudinal displacement of the particles, which enters into oscillation in equilibrium is represented by a sinusoidal function:

$$x = A \sin \omega t .$$

Where: A – the amplitude (the particle maximum movement from the equilibrium position); ω – the cyclic frequency ($\omega = 2\pi\nu$); $T = \frac{1}{\nu}$ – the period.

The distance traveled by the sound in an environment, during a period of time is called *wavelength* (λ) :

$$\lambda = cT = \frac{c}{\nu} .$$

Where c – the velocity of sound in the given environment.

It should be mentioned that the frequency (ν) is specific for sound sourse and it is independent of the environment, while the wavelength (λ), depends on the sound velocity, and varies on different environments.

Two pure sounds may impose particular longitudinal displacement of a common environment, but these movements can be triggered in time.

Between $X_1 = A\sin\omega t$ and $X_2 = A\sin(\omega t + \varphi)$ there is no difference in the movement of particles, but the second movement, in relation to the first, is in advance of time:

$$\tau = \frac{\varphi}{\omega}.$$

Where: $(\omega t + \varphi)$ – phase oscillations; φ – initial phase.

Instantaneous speed of the particles (V) is the first derivative according to time of the displacement function:

$$x = A\sin\omega t.$$

$$V = \frac{dx}{dt} = A\omega\cos\omega t.$$

Complex sounds and noises

Complex sounds impose periodic movements to environmental particles, but they are not sinusoidal.

Based on Fourier's theory, **any periodic motion can be considered as a superposition of a finite or infinite number of sinusoidal movements of various amplitudes, with frequencies that are multiple of the fundamental frequency** (*fig. 20.2*).

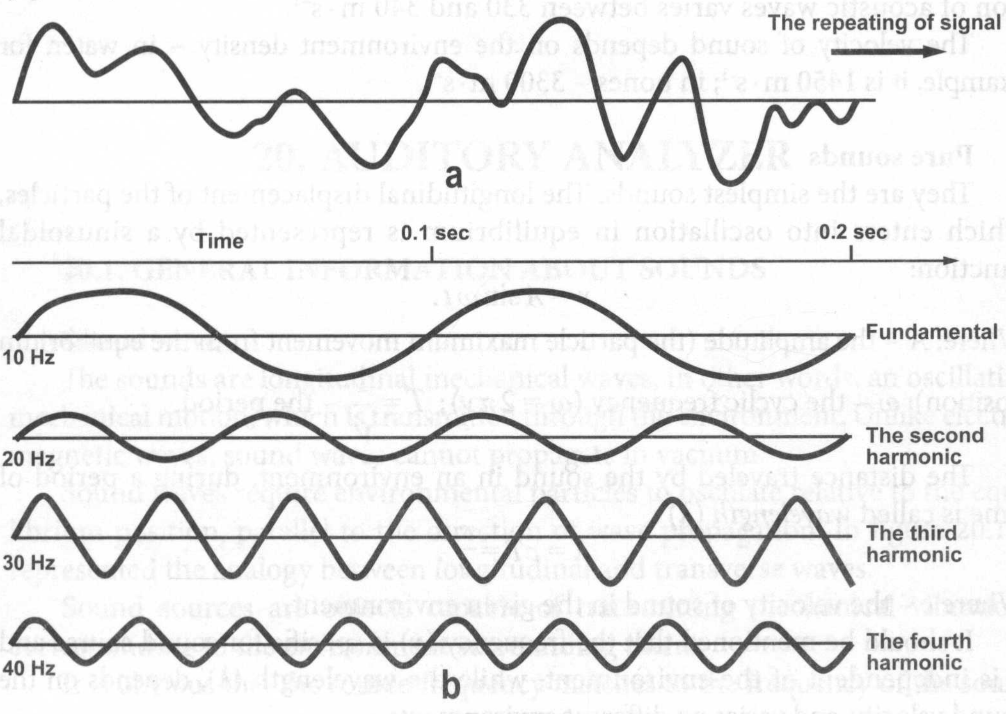


Fig. 20.2.

In a complex sound can be distinguished:

- A non-sinusoidal periodic oscillatory motion;
- Fundamental sound frequency **ν**;
- A certain number of harmonics with higher frequencies, multiple frequency ν.

$$X_{re} = A_1\sin(\omega\,t+\varphi_1)+A_2\sin(2\,\omega\,t+\varphi_2)+\ldots+A_n\sin(n\omega+\varphi_n).$$

The amplitude-frequency characteristics of all harmonic components represent the **harmonic spectrum** of complex sound.

In figure 20.3 is represented the harmonic spectra of acoustic signals emitted by different musical instruments.

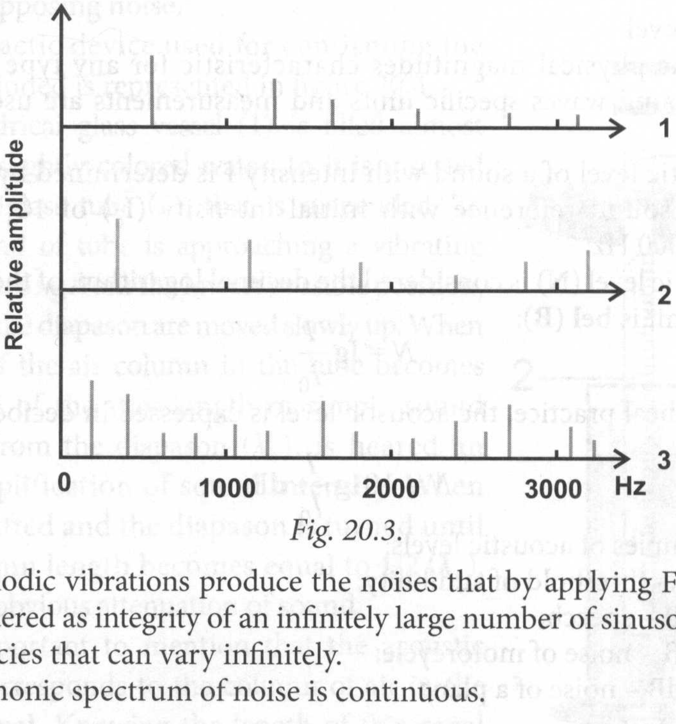


Fig. 20.3.

Non-periodic vibrations produce the noises that by applying Fourier's theory can be considered as integrity of an infinitely large number of sinusoidal functions, with frequencies that can vary infinitely.

The harmonic spectrum of noise is continuous.

Sound pressure (P) is determined from the relationship:

$$\mathbf{P = v\rho c}.$$

Where: ρ – the density of the enviroument; c – velocity of sound in the given environment; v – instantaneous velocity of the particles included in vibration.

Acoustic intensity

The energy carried by the sound wave through a unit of area per unit time represents the **sound intensity**; it is determined, for pure sound by the relationship:

$$I = vP; \quad P = v\rho c; \quad I = v^2\rho c.$$

For sinusoidal function: $v^2 = \frac{1}{2}\omega^2 A^2$.

The final expresion for medium value of acoustic intensity is:

$$I = \frac{1}{2}\rho c \omega^2 A^2.$$

The link between sound intensity and acoustic pressure is determined from:

$$I = \frac{P^2}{\rho c}.$$

Where: ρ – the enviroument density in which sound propagates, c – sound velocity.

Acoustic level

Besides the physical magnitudes characteristic for any type of waves, to characterize sound waves specific units and measurements are used, including **acoustic level.**

The acoustic level of a sound with intensity I is determined by comparison with an pure sound reference with initial intensity (I_0) of 10^{-12} W/m^2 and frequency of 1000 Hz.

The acoustic level (N) is considered the decimal logarithm, of the relation $\frac{I}{I_0}$, the measure unit is **bel** (**B**):

$$N = \lg\frac{I}{I_0}.$$

In the medical practice, the acoustic level is expressed in decibels (**dB**) from the relation:

$$N = 10\lg\frac{I}{I_0}\ \text{dB}.$$

Some exemples of acoustic levels:

0 dB – threshold of audibility;
20 dB – speech;
90 dB – noise of motorcycle;
130 dB – noise of a plane.

The acoustic impedance represents an important characteristic of the enviroment. In one point of enviroument, the acoustic impedance is equal with the report between acoustic pressure and vibration of the sound (v) of particles in this point:

$$Z = \frac{P}{v} = \frac{v \cdot \rho \cdot c}{v} = \rho c.$$

Some values of impedance in different environments:

Air z = $0.04 \cdot 10^3$ g·cm^{-2}·s^{-1};
Water z = $1.48 \cdot 10^5$ g·cm^{-2}·s^{-1};
Bone z = $7.80 \cdot 10^5$ g·cm^{-2}·s^{-1}.

Acoustic resonance

When the sound propagates freely in a limited medium and its frequency coincides with the frequency of the medium, the oscilation amplitude increases; this fenomen is called **acoustic resonance**. An example of acoustic resonators can be columns of air, which fill cavities with several sizes and shapes.

When in resonator penetrates a complex sound wave, some of harmonic tones (with multiple frequencies of its own frequencies) are amplify, and other (with non-multiple frequencies of its own frequencies) are attenuated.

In the second case, the resonator serves as a **damper** of sounds – device used usually for opposing noise.

The didactic device used for confirming the effects mentioned is represented in figure 20.4.

A cylindrical glass vessel (1) is filled almost to top with slightly colored water. In it is inserted a graduated glass tube (2), that is suspended. To the upper end of tube is approaching a vibrating diapason. Starting from the lowest possible position, the tube and the diapason are moved slowly up. When the length of the air column in the tube becomes equal to 1/4 of the wavelength of simple sound transmited from the diapason (λ_s), is heared an obvious amplification of sound intensity. When the tube is lifted and the diapason is turned until the air column length becomes equal to 1/2 (λ_s), it occurs an obvious attenuation of sound.

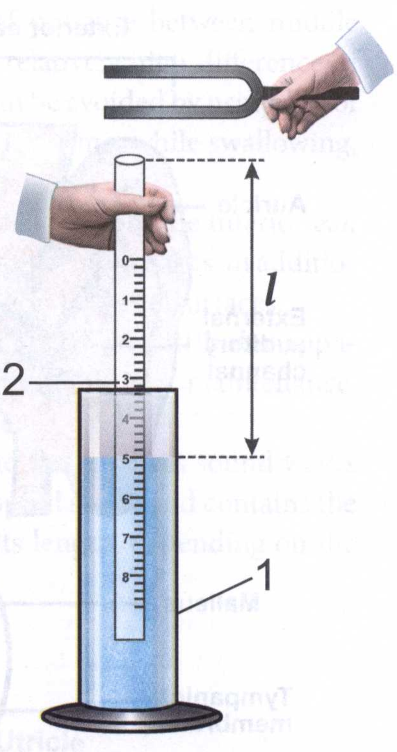


Fig. 20.4.

It is important to mention that the acoustic resonator corresponds to the column of air in the **auditory canal**. Knowing the length of this canal ($l \approx 2.3$ cm) and taking in to consideration the results of the experience described, we can easily determine the simple sound frequency, which amplifies to the maximum in the auditory canal:

$$\nu = \frac{c}{\lambda} = \frac{3.5 \cdot 10^2}{4 \cdot 2.3 \cdot 10^{-2}} \approx 3.5 \text{ kHz}.$$

20.2. ANATOMY AND PHYSIOLOGY OF THE EAR

The auditory analyzer establishes a direct connection of the receiver of sound waves with the central nervous system. Using the notion of cybernetics, we can say that auditory analyzer receives and transmits information.

In the following chapter, from the whole auditory system we will mention the structure and physical processes which occur in the ear.

Ear is the organ used for detection and precautionary processing of sound. It consists of three sections: **exterior ear**, **middle ear** and **interior ear** (*fig. 20.5*).

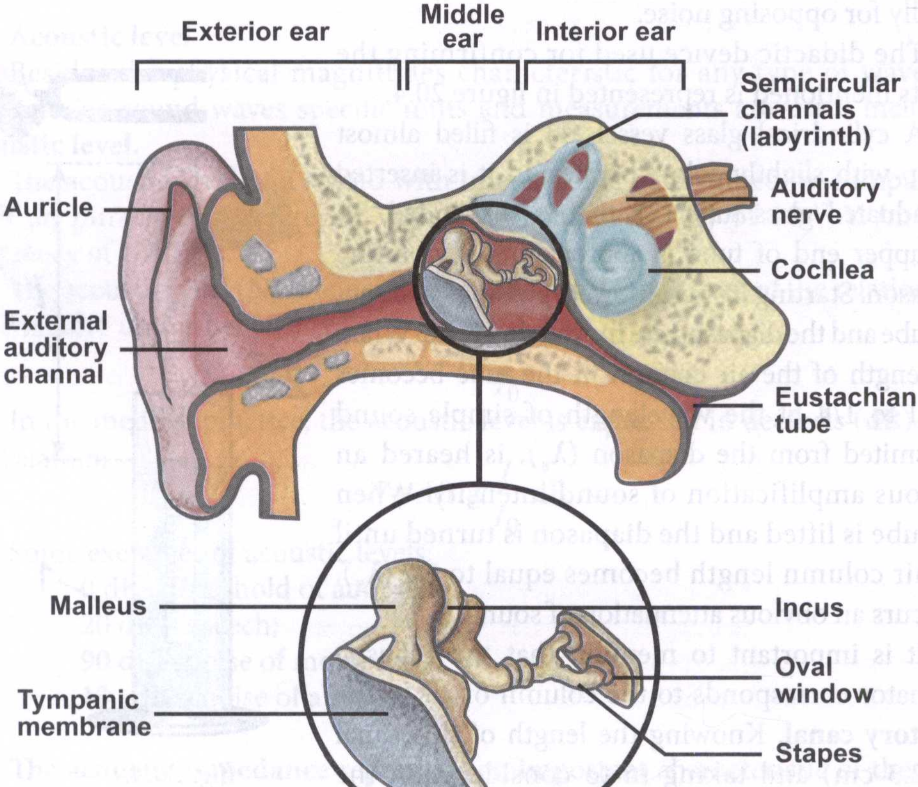


Fig. 20.5.

The exterior ear is formed of an **auricle**, which captures incident sound waves, **auditory canal**, which leads the sound energy to the middle ear.

In the auditory canal occurs **first amplification** of sound due to the **resonance phenomenon**. Passing through the auditory canal, sound waves put in oscillation the tympanic membrane, located at the entrance of the middle ear.

Middle ear is a cavity (15 × 5 × 2 mm) containing three ossicles, each of them with a characteristic shape: **hammer**, **anvil** and **ladder**.

Tympanic membrane vibrations are transmitted firstly to the hammer, then to the anvil and ladder, and from here, to the **oval window**, *situated at the boundary between middle ear and the interior ear one.*

The ossicles work like a leverage that decreases the range of motion, increasing the power. Thereby is achieved **the second amplification** of the sound.

The middle ear's space communicates with the nasal cavity through the **Eustachian tube**. Normally this communication is closed. If one goes in a high altitude or dives into water, there is a difference of pressure between middle ear and the environment. This causes pain and at a relatively high difference of pressure the tympanum is in rosk of rupturing. This can be avoided by using one of *Eustachian tube's* functions, which is opening for a short time, while swallowing, and so the pressures are equilibrated.

Oval window and **round window** separate middle ear from the interior ear. The oval window is 14–20 times smaller than tympanum, which makes an additional amplification of acoustic pressure, equal to the ration between surfaces.

The inner ear includes the organ of hearing (**cochlea**) and **vestibular apparatus**, which is an organ of spatial orientation and equilibrium maintenance, consisting of **three semicircular canals** and **vestibule.**

Cochlea is a cavity filled with endolymph – liquid that receives sound waves sent through external and middle ear. Cochlea has a spiral shape and contains the basilar membrane, whose resonance is different on its length, depending on the frequency of the sound waves.

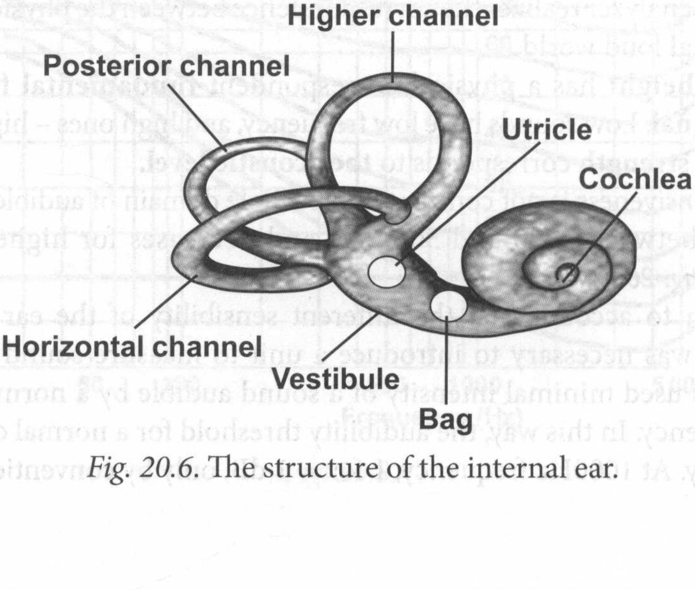


Fig. 20.6. The structure of the internal ear.

In other words, in the interior ear occurs a Fourier analysis of the received sounds. At the surface of the basilar membrane there is a cellular layer known as **the organ of Corti**, consisting of **sensory cells (cilia)**. The vibrations that are passing through the cochlear channel put in motion the **basilar** and **tectorial membranes**, which act on the cilia. Inside these sensorial cells takes place the **encoding of mechanic oscillations in action electrical potentials** sent to the brain, where they are transformed into **auditory sensations**.

From what was said above, it obeys that sound processing in auditory analyzer occurs in two phases:

1) the capture and the processing of sounds maintaining the nature of mechanical oscillations;

2) mechanical oscillations encoding in action electrical biopotentials to sensorial cells, then their neuronal processing which lead to auditory sensation.

The spatial orientation and equilibrium maintenance is accomplished with the aid of **membranous labyrinth**, **semicircular canals** with an essential role. Gravity and acceleration modifications of the body determine changes in fluid dynamics in the three semicircular canals, which in their turn, act on cilia of the sensitive cells present both in semicircular canals and in utricle and saccula. Then, the information is sent through vestibular nerve, to the cerebellum, which transform them in awareness about the position of the body towards the direction of gravitational acceleration and then in decisions to maintain the balance.

20.3. SUBJECTIVE PHENOMENA OF AUDITORY ANALYZER

Auditory analyzer realizes the correspondence between the physical acoustical and **perceptual** loud world.

1. **Tonal height** has a physical correspondent **fundamental frequency of acoustical signal**. Low sounds have low frequency, and high ones – high frequency.

2. **Sound strength** corresponds to **the acoustic level.**

Ear responsiveness is not constant in the whole domain of audible frequencies. It is maxim between 2000 and 5000 Hz and decreases for higher and lower frequencies (*fig. 20.7*).

Taking in to account that the different sensibility of the ear depends on frequency, it was necessary to introduce a unit to measure sound strength. As reference it is used minimal intensity of a sound audible by a normal ear at respective frequency. In this way, the audibility threshold for a normal ear is **0 fon** at any frequency. At 1000Hz frequency, 1 fon = 1 dB, only by convention.

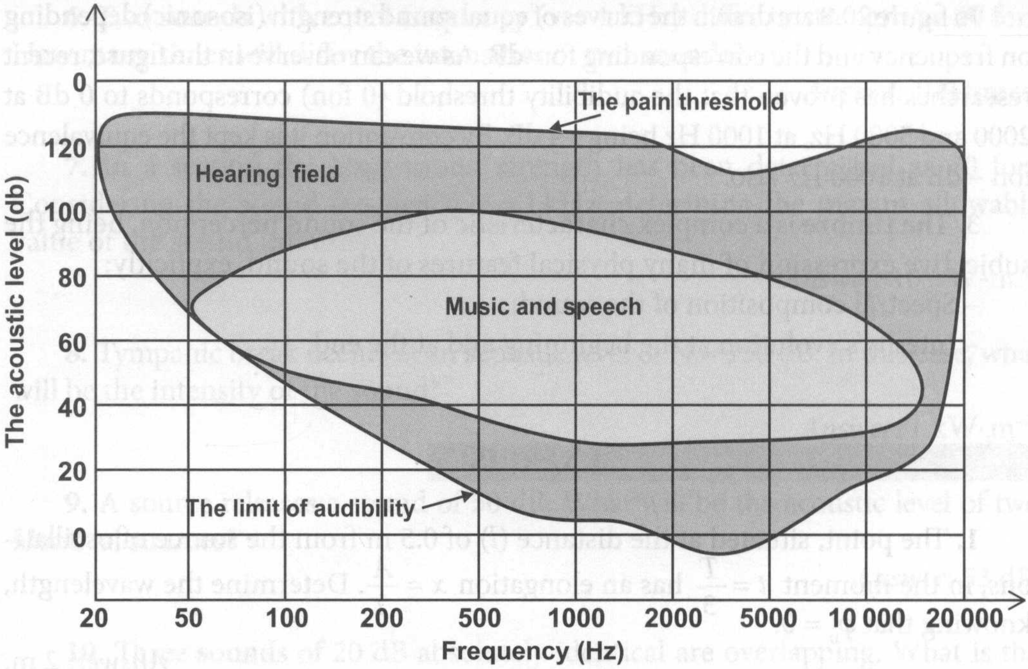


Fig. 20.7.

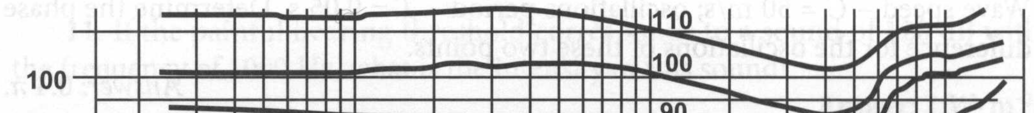


Fig. 20.8.

In figure 20.8 are drawn the curves of equal sound strength (isosonic) depending on frequency and the corresponding fon-dB. As we can observe in the figure, recent researches has proven that the audibility threshold (0 fon) corresponds to 0 dB at 2000 and 5000 Hz, at 1000 Hz being +4 dB. By convention it is kept the equivalence fon – dB at 1000 Hz also.

3. **The timbre** is a complex characteristic of the sound perception, being the subjective expression of many physical features of the sound, explicitly:

– Spectral composition of the sound;

– Intesity's evolution at the beginning and at the end.

EXERCISES

1. The point, situated at the distance (l) of 0.5 m from the source of oscillations, in the moment $t = \frac{T}{3}$ has an elongation $x = \frac{A}{2}$. Determine the wavelength, knowing that $\varphi_0 = 0$.

Answer: 2 m.

2. Distance difference of two points from the wave source (Δl) is 50 cm. Wave speed – $C = 50$ m/s; oscillations period – $T = 0.05$ s. Determine the phase difference for the oscillations of these two points.

Answer: 0.4 π.

3. Route difference of sound waves which come into the ear (Δl) is 1 cm. Determine the difference of phase between both auditory sensations for the tone with frequency $\nu = 1000$ Hz.

Answer: $\frac{\pi}{17}$.

4. Determine the average value of the force, which acts on human tympanic membrane ($S = 66$ mm^2) in case of pain, at the threshold for sound frequency of 1000 Hz.

Answer: ≈ 6.17 mN.

5. How much will increase the sound strength if its intensity increased 1000 times from the threshold? Solve the exercise for sound with frequency: a) 100 Hz; b) 1000 Hz.

Answer: with 45 and 30 fon.

6. Two sounds with equal frequency ($\nu = 1$ kHz) differ in strength by 20 fon. How many times will differ the intensities of two sounds?

Answer: 100 times.

7. In a section the limit sound strength has been determined as 70 fon. Considering the sound frequency $\nu = 1\,\text{kHz}$, determine the maxim allowable value of the sound intensity.

Answer: $10^{-5}\ \text{W}\cdot\text{m}^{-2}$.

8. Tympanic break occurs at an acoustic level of N = 150 dB. In this case, what will be the intensity of the sound?

Answer: $1\ \text{kW}\cdot\text{m}^{-2}$.

9. A source releases a sound of 50 dB. What will be the acoustic level of two identical sources?

Answer: 53 dB.

10. Three sounds of 20 dB absolutely identical are overlapping. What is the acoustical level of the resulting sound?

Answer: 24.77 dB.

11. If the painful hearing threshold corresponds to a sound of 120 dB with the frequency of 1000 Hz, what is the intensity of this sound?

Answer: $1\ \text{W}\cdot\text{m}^{-2}$.

12. A sound with 1000 Hz frequency and intensity (I_1) of $10^{-7}\ \text{W}\cdot\text{m}^{-2}$ is issued simultaneously with a sound of 2000 Hz frequency and intensity (I_2) of $10^{-4}\ \text{W}\cdot\text{m}^{-2}$. Determine the acoustical level of the resulting sound.

Answer: $\approx$ 80 dB.

13. Let's consider two sounds with the same acoustical intensity: $I = 10^{-6}\ \text{W}\cdot\text{cm}^{-2}$. One has frequency of 200 Hz and the other – de 600 Hz. What will be the acoustic level at the simultaneously emission of these sounds?

Answer: $\approx$ 103 dB.

14. Two sounds of same frequency differ by acoustical level by 30 dB. Determine the ratio between the amplitudes of acoustical pressure.

Answer: $\dfrac{\rho_1}{\rho_2} = 32.$

21. VISUAL ANALYZER

21.1. EYE AS AN OPTICAL INSTRUMENT

Visual analyzer is a biological system, (sensitive to electromagnetic radiation in the wavelength 400–750 nm) by which the body receives information about the shape, size, position, motion, brightness and color of objects ambience. At the same time, it is an organ of expression of mental states (transmitter of information).

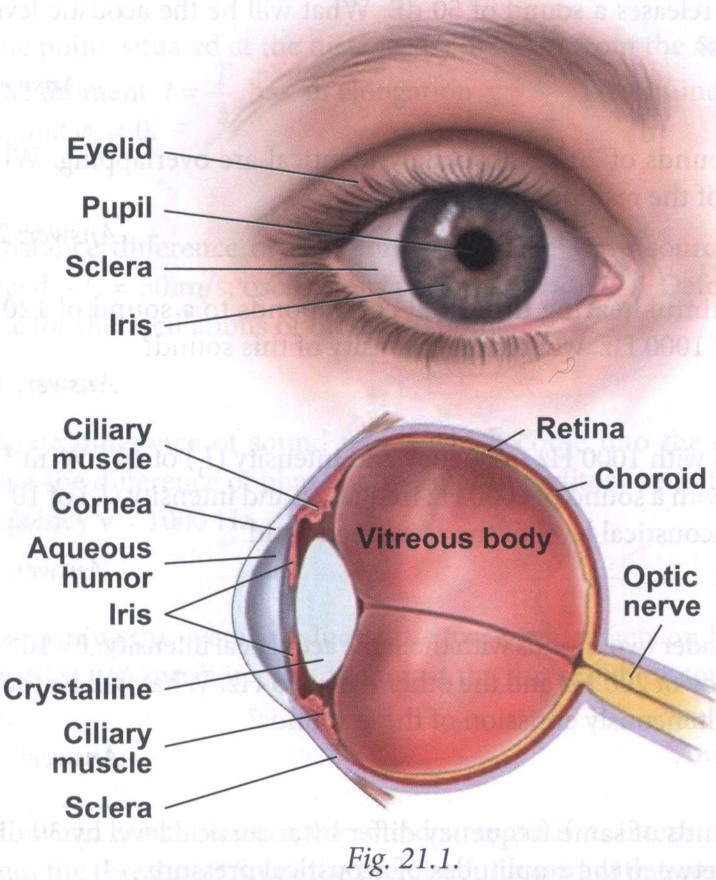


Fig. 21.1.

The eye, is the center of the optical system (*fig. 21.1*), consisting of four transparent media: **cornea, aqueous humor**, **lens** and **vitreous humor** with different refractive index (**n**): cornea and aqueous humor, n = 1.337, the lens has a refractive index that varies between 1.375 and 1.473, and for vitreous humor n = 1.336.

It is known that optical power (converging) in a spherical lens is the inverse size of the focal length:

$$D = \pm \frac{1}{f}\, m^{-1} \text{ (dioptres).}$$

For determination of the convergence of a spherical diopter (*fig. 21.2*) with refractive index (**n**) located in air (**n** = 1) is of great importance due to the fact that such diopter is **the cornea** of the eye. For this purpose we use the equation:

$$D = \frac{n-1}{R}.$$

Where $R = \overline{SC}$ is the radius of curvature of the diopter, expressed in meters.

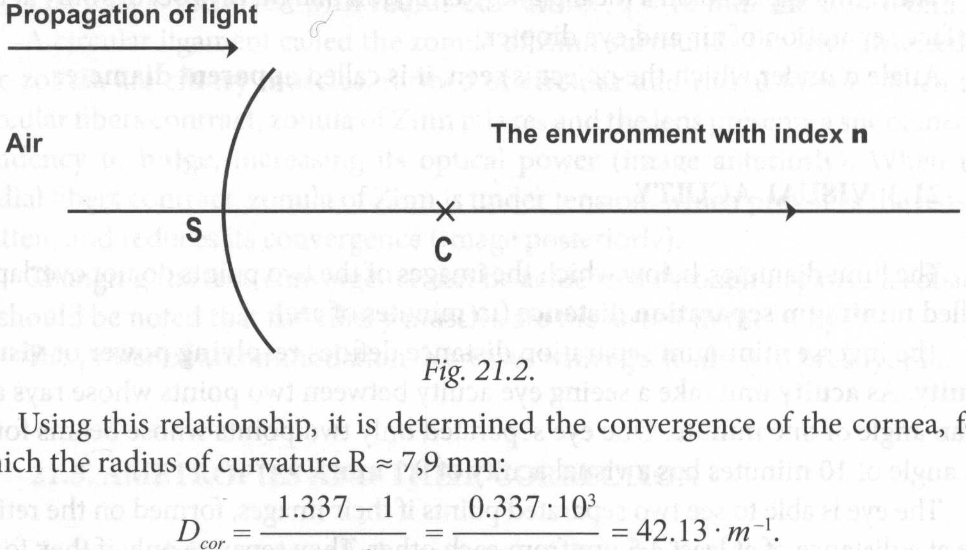


Fig. 21.2.

Using this relationship, it is determined the convergence of the cornea, for which the radius of curvature R ≈ 7,9 mm:

$$D_{cor} = \frac{1.337-1}{7.9 \cdot 10^{-3}\, m} = \frac{0.337 \cdot 10^{3}}{7.9\, m} = 42.13 \cdot m^{-1}.$$

21.2. REDUCED EYE

In various researches a simplified model, called the reduced eye, replaces **normal eye**. In it, all spherical dioptres are represented by a single one (*fig. 21.3*).

Below, it is represented a simplified optical diagram of the eye that allows to study the functioning normal eye, called **emmetropic**, however an abnormal one is called **ametropic**, referring to the eye as an optical system.

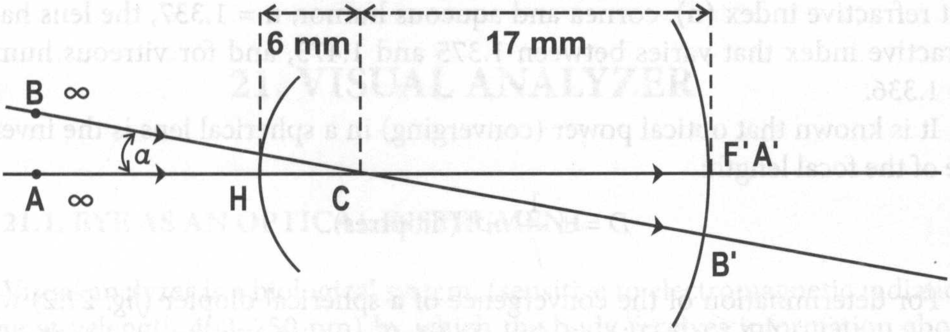


Fig. 21.3.

Convergence system is about 60 diopters. The distance between the optical center and the retina is about 17 mm, and the eye diameter is approximately 23 mm.

The retina is placed on a focal plane. Refraction of light rays occurs only at the surface separation of air and eye diopter.

Angle α under which the object is seen, it is called **apparent diameter**.

21.3. VISUAL ACUITY

The limit diameter, below which the images of the two points do not overlap is called **minimum separation distance** (in minutes of arc).

The inverse minimum separation distance defines **resolving power** or **visual acuity**. As **acuity** unit take a seeing eye acuity between two points whose rays are at an angle of one minute. One eye separated only two points whose beams form an angle of 10 minutes has a visual acuity of 0.1 min^{-1}.

The eye is able to see two separated points if their images, formed on the retina are at a distance of at least 4.5 μm from each other. They separate only if they form two different elements of the retina, with the size. Calculations show that in reality the distance between the two points must be at least 60 mm, when they are at a minimum distance of distinct vision (25 cm).

Visual acuity changes with age: its maximum is reached at 14 years in boys and girls at 12 years. Besides age, visual acuity depends on:

– diopter factors such as spherical and chromatic aberration, ametropies, pupillary diameter and others;

– retinal factors, which bind to discontinuous granular structure of the retina that is the center spot image, must lie on distinct receptor cells;

– factors on the system: in detail, contrast, brightness, exposure, composition, colors (monochromatic light increases visual acuity by eliminating chromatic aberration).

To increase the separation power of the eye, you have to help it with various optical instruments.

Visual acuity is called **amblyopic**.

21.4. ACCOMMODATING TO FORM CLEAR IMAGES

For an image to be perceived clearly, it must be formed on the retina. **Accommodating** for formation of sharp images is achieved by **modifying the crystalline curvature.**

The lens is biconvex with radius curvature: $R_1 = 10$ mm and $R_2 = 6$ mm.

A circular ligament called the zonula of Zinn surrounds the lens. Inserted in the zonula are **ciliary muscles**, formed of **circular** and **radial fibers**. When the circular fibers contract, zonula of Zinn relaxes and the lens presents a spontaneous tendency to **bulge**, increasing its optical power (image anteriorly). When the radial fibers contract, zonula of Zinn is under tension, which provokes the lens to flatten, and reduces its convergence (image posteriorly).

Changing the lens convergence can be achieveed through negative feedback. It should be noted that the ciliary muscles are the fastest in the body.

The power of accommodation decreases with age, leading to **presbyopia.**

21.5. AMETROPIES AND THEIR CORRECTION

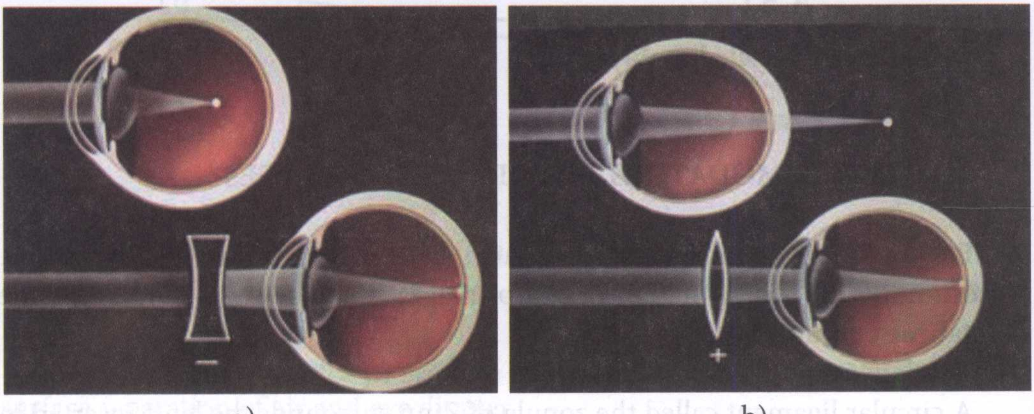
Imaging defects are called **ametropies. Punctum proximum** (m) is the nearest point on the main optical axis, in which an object is clearly seen with high accommodation. Punctum remotum (M) is the farthest point in which the object can be seen clearly, without adjustment. For emmetropic eye m = 25 cm and $M = \infty$.

Myopia

Punctum **proximum** and punctum **remotum** are smaller than normal (*fig. 21.4 a*). **Axial myopia** is very common. In this condition, the anterior-posterior axis of the eye is longer than normal.

Curvature myopia. The curvature of the lens, most commonly the posterior is greater, due to eye fatigue.

Index myopia. The refractive index increases due to large concentration of saline in the refractive medium.



a) b)

Fig. 21.4.

Hyperopia

Punctum **proximum** and punctum **remotorum** are higher than normal. The image is formed behind the retina (*fig. 21.4 b*).

Axial hyperopia. Anterior-posterior axis is shorter.

Hyperopia curvature. The lens is more elongated, so it has lower convergence.

Hyperopia index: n decreases due to decrease in salt concentration. It's more an abstraction than a practical reality.

Presbyopia – hyperopia old (reducing capacity of accommodation).

Astigmatism

Occurs when the radius of curvature, and thus focal lengths are not equal from one meridian to another in the eye diopter (especially in the cornea).

In the case when the radius of curvature varies from one meridian to another and is continuous, is called **regular astigmatism** (*fig. 21.5*).

Irregular astigmatism occurs only when the radius of curvature varies irregularly and accidentally from one meridian over another. Astigmatism is often accompanied by another ametropia.

Ametropies correction is made by:

- **diverging** lens for **myopia;**
- **converging** lens for **hyperopia;**
- **cylindrical** lens for **regular astigmatism**.

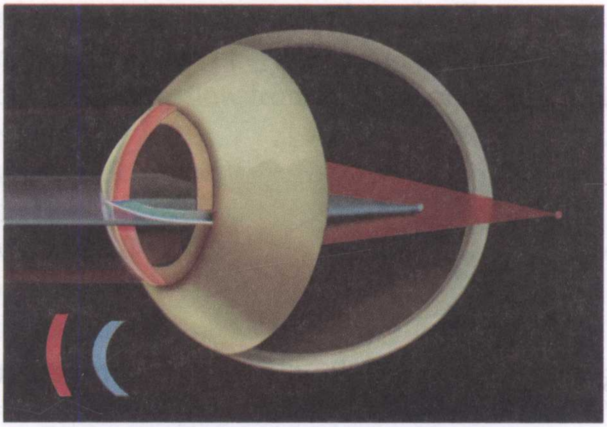


Fig. 21.5.

Irregular astigmatism is a very complicated problem: until recently was corrected with **individual contact lenses**. It can be treated with laser radiation, removing an elliptical surface in a particular meridian.

Cataracts

Is the change in lens transparency. Require surgery.

Using lasers to correct ametropies

The first intervention on the human eye to radiation was performed in 1988 in Germany and was called **PRK** (keratectomie photoreactive). This method is perfected continuously for correcting myopia, hyperopia and astigmatism. Another technique called **LASIC** (*laser assisted in situ keratomileusis*) is more effective in severe refraction. Being completely painless has a dramatic effect: a few hours after surgery, the patient is fully recovered. In essence, during a surgery, laser beam, guided by computer, polishes cornea's curvatureand modelates it depending on the type and degree of refractive error. With great precision, the laser removes tissue and ultra thin layers. In case of myopia, for example, laser eye axis shortens, flattens the cornea, in one of hyperopia, laser lengthens it, tissue is removed from the periphery. In regular astigmatism, an elliptical surface in a given meridian of the cornea is removed.

Laser therapy, used in retina is CO_2 laser with transparent media of the eye that can travel without being absorbed by it; the whole energy of the retina sticks to the sclera with photocoagulation. The laser is used in the treatment of glaucoma, allowing restoration of intraocular fluid drainage system and thus reducing intraocular pressure (*fig. 21.6*).

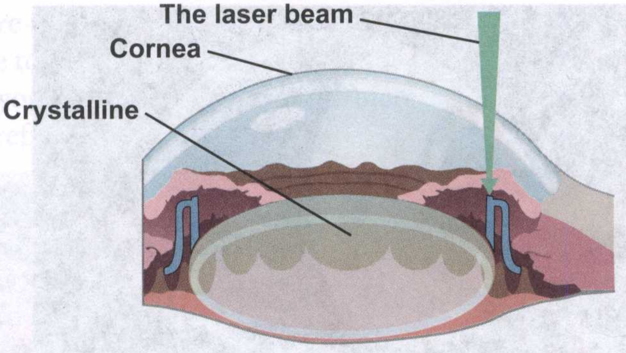


Fig. 21.6. Laser surgery to restore the drainage system.

21.6. BIOPHYSICAL BASES OF VISUAL SENSATION. THE MECHANISM OF COLOR VISION

The structure of the retina

Retina has a thickness of 350 μm and consists of six types of cells, arranged in successive structures, each with a specific role (*fig. 21.7*). Latest in the path of light are **epithelial cells pigment** followed by **rod** and **cone photoreceptor cells** which contain **photosensitive pigments.**

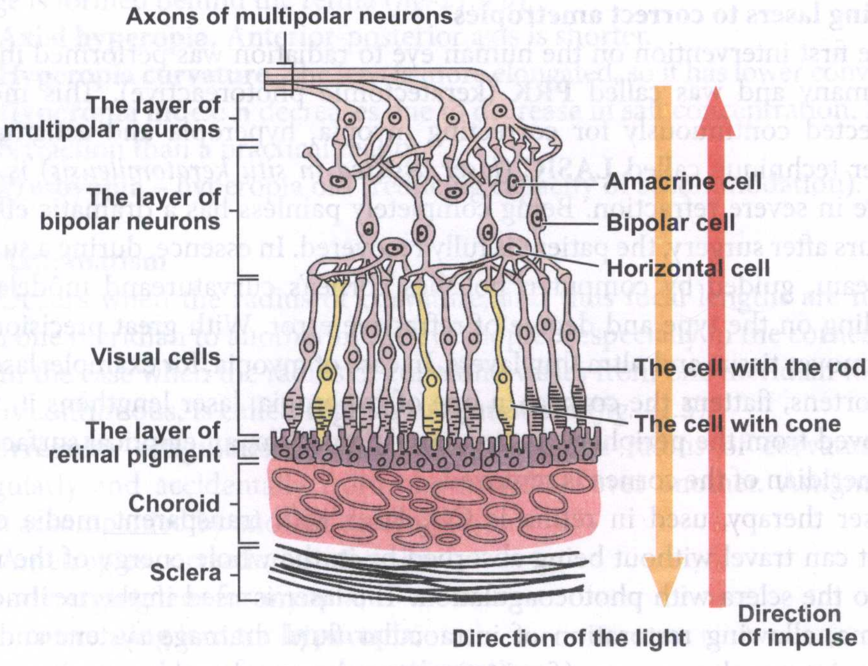


Fig. 21.7.

The sensitivity of the eye

Metering is to assess the level of brightness of eyes like receptor selective. Eye reaction varies not only energy flow from the source of light, but the wavelength of the radiation.

Light-sensitive retinal cells, **rods** and **cones**, fulfill different roles in achieving the sensation of light. Rods are more sensitive to light, but do not distinguish colors. They provide **scotopic** vision (gr. **scotos** – *dark*) in low light colorless (black and white).

Cones provide **photopic** vision (gr. **fotos** – *light*) to bright light and color vision; they have a lower sensitivity than rods.

At a sufficient brightness of the lens, cones are sensitive to the assessment of image. Therefore, the resolution power of the eye depends on the distribution of cones on the retina.

Inverse brightness treshold (minimum brightness which causes visual sensation) is called **sensitivity of the eye**.

Eye sensitivity (varies widely due to visual adaptation) is achieved through several ways: by varying the diameter of the pupil, by reducing light sensitive pigment concentration, the shielding cones and rods by a dark pigment, by varying the brightness of the object and the degree of participation of cells responsive to visual sensation arousal.

Adaptation allows the eye to function normally in the range of brightness from 10^{-7} until 10^5 *lx*.

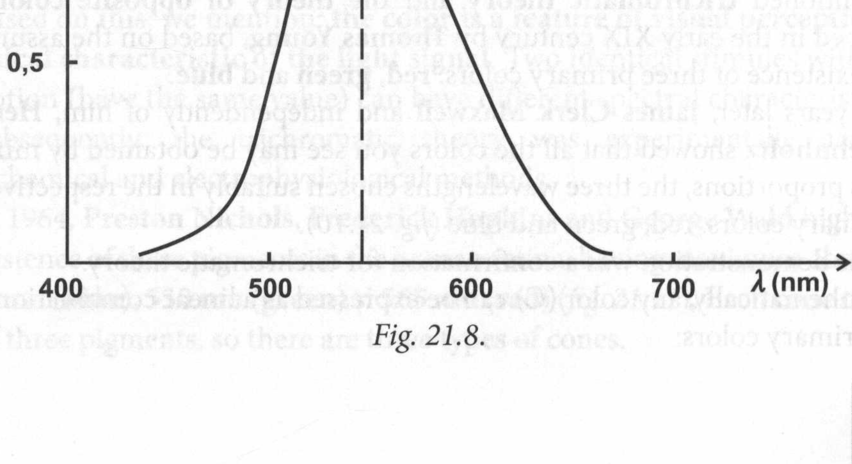


Fig. 21.8.

Spectral sensitivity of the eye is characterized by the coefficient of efficiency of light (V_λ), which ranges from 0, for wavelengths invisible to 1, wavelength **λ_m = 555 nm**, where the eye sensitivity is high (*fig. 21.8*).

Maximum photopic vision sensitivity curve corresponds to the maximum spectral density of solar radiation that passes through the atmosphere, reaches the surface of the Earth. This manifests rationality and structurs the human eye.

In order to scotopic, maximum sensitivity curve in relation to photopic vision is gone and corresponding wavelength **λ = 510 nm** (*fig. 21.9*).

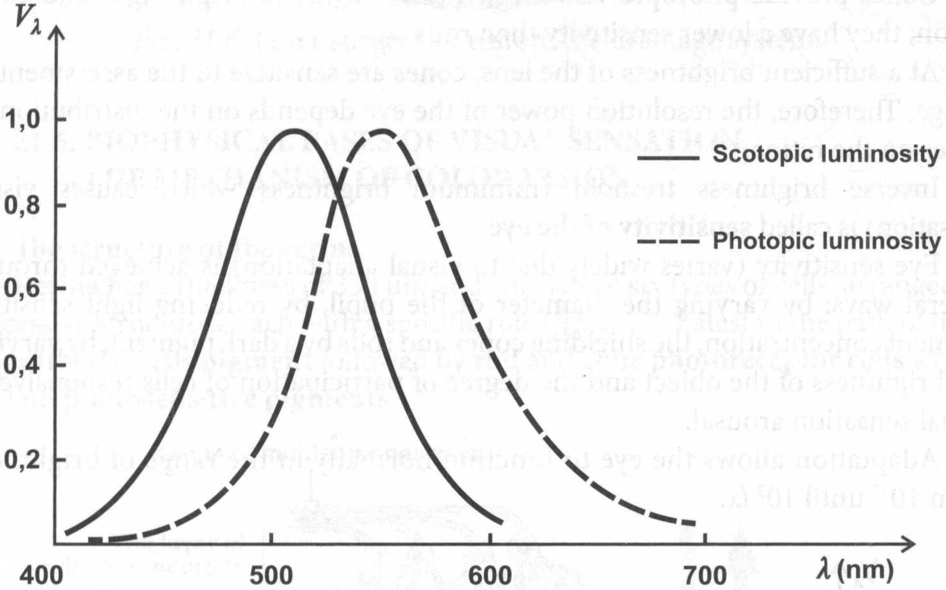


Fig. 21.9.

Mechanism of color vision

With reference to color vision were developed several theories. Among them are mentioned **trichromatic theory** and **the theory of opposite color**. First developed in the early XIX century by **Thomas Young**, based on the assumption of the existence of three primary colors: **red**, **green** and **blue**.

50 years later, **James Clerk Maxwell** and independently of him, **Hermann von Helmholtz** showed that all the colors you see may be obtained by mixing in various proportions, the three wavelengths chosen suitably in the respective fields the primary colors: red, green and blue (*fig. 21.10*).

This demonstration was a confirmation for trichromatic theory.

Mathematically, any color (**C**) can be expressed as a linear combination of the three primary colors:

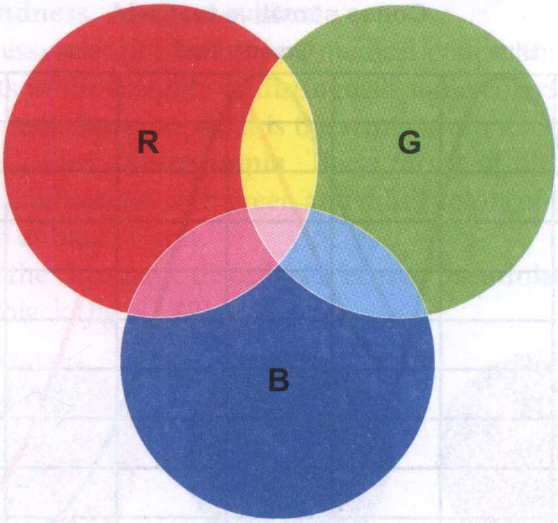


Fig. 21.10.

R (red), G (green) şi B (blue):

$$C = xR + yG + zB.$$

Where: **x**, **y** and **z** are called **chromatic coefficients** and represent the primary color in color C. They can take positive or null. A certain sense of color can be obtained from the combination of different wavelengths of radiation. Example: the mixture of yellow and blue radiation reproduces green color.

There are certain rules for combining colors, namely:

– if two signals, C and C^I have the same color ($C = C^I$), then overlap them a third signal (C^{II}), we obtain also identical colors ($C + C^{II} = C^I + C^{II}$);

– if it increases the intensity of each primary color in the same report, chromatic coefficients multiply the same number of times (**nC, nC^I, nC^{II}**), the resulting color is the same as the original, but with a brightness (intensity) higher.

Based on this, we mention: **the color is a feature of visual perception, not a physical characteristic of the light signal**. Two identical stimules with visual perception (have the same value) can have different spectral characteristics.

Subsequently, the trichromatic theory was experimentally according photochemical and electrophysiological methods.

In 1964, **Preston Nichols**, **Frederick Hopkins** and **George Wald** highlighted the existence of three pigments in the human retina, having maximum absorption at **420 nm (blue)**, **530 nm (green)** şi **565 nm (red)** (*fig. 21.11*). Each cone contains one of three pigments, so there are three types of cones.

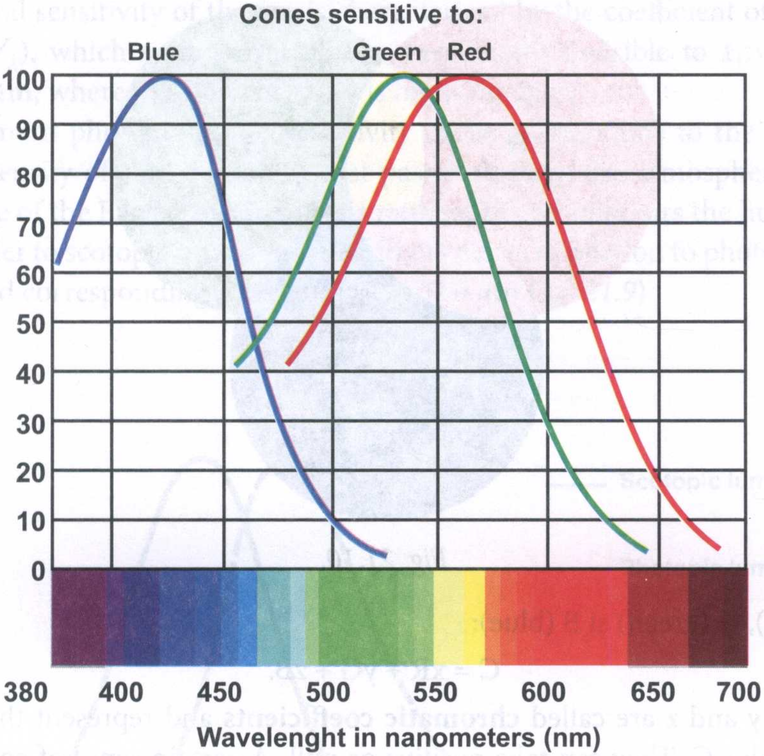


Fig. 21.11.

Black gives the feeling of a lack of light, but this note is absolutely necessary. Therefore, the sensation of black is linked to a specific retinal activity, possibly under the influence of invisible radiation.

The second theory, mentioned in the beginning belongs to **Eswald Hering** and is less recognized. According to this theory, there are three pairs of opposite colors: red and green, yellow and blue, white and black. In the retina, for each pair of colors are special chemicals. Under the action of light, there is destruction of these substances **(dissimilation)** and the dark-restore **(assimilation)**. Different combinations of dissimilation and assimilation processes create sensations of color diversity.

The results obtained by the most modern methods confirmed that **the trichromatic theory** is valid in the photoreceptors, and **the opposite color** – in transmitting information.

The color blindness. Medical context

Color blindness, scientific name, is a medical condition; a disorder of color vision that consists in the inability to distinguish some colors from others (especially **red** from **green**). **Protanopia**, it is the name given to blindness for red and green blindness is called **deuteranopia**. These forms of blindness express that those people even confuse red with green and other colors of the same brightness or saturation, such as blue or gray.

To investigate the chromatic disorders it is used **pseudoisocromatics** boards, Ishihara or atlas Polack (*fig. 21.12*).

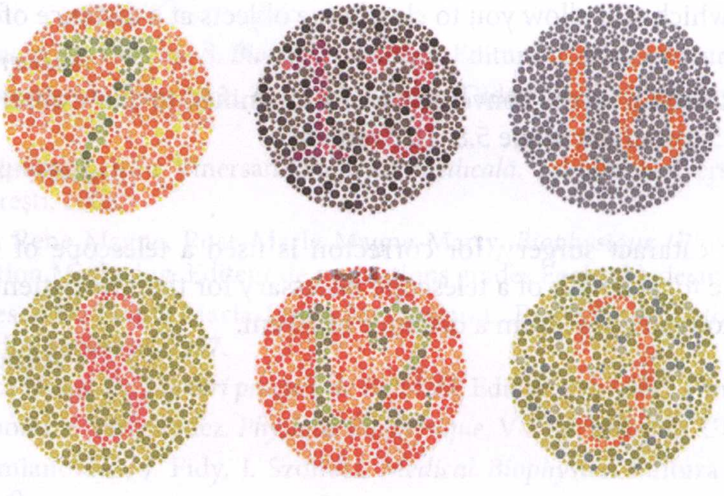


Fig. 21.12.

These samples consist of drawings that are presented by circles of different sizes, colors, shades, numbers, letters or figures. The subject placed on a plane, is required to identify numbers, letters or printed sign. This plan shall be such, so that the normal eye can read them clearly, while those who are color blind identify numbers, letters or symbols with weight incorrectly or not at all. Color blindness is a congenital disease caused by a failure of a portion of the retina or optic nerve. Although the disease is inherited maternally, especially men suffer from it.

EXERCICES

1. Upper limit of the flow of light entering the human eye that does not cause pain is $2 \cdot 10^{-5}$ W. Determine the number of photons entering the eye in this case, if the wavelength of radiation is 555 nm radiation.

Answer: $55.8 \cdot 10^{12}$ fotons.

2. Visual sensation at human occurs when the energy of bright radiation entering the eye is $2 \cdot 10^{-13}$ J. How many photons with a wavelength of 700 nm have to penetrate together in the eye to produce visual sensation?

Answer: $0.7 \cdot 10^6$ fotons.

3. At what distance from the object should be placed a lens with focal length f = 10 cm to get a magnification equal to 10?

Answer: 9 cm.

4. An eye has proximum point located at 1.5 m. Determine the optical power of the lens, which will allow you to clearly see objects at a distance of 25 cm.

Answer: $+3.33$ m^{-1}.

5. What should be the convergence of a magnifier, to allow observation of an object of 12.5 μm under angle $5.8 \cdot 10^{-4}$ radians?

Answer: 46.4 diopters.

6. After cataract surgery, for correcton is used a telescope of 15 diopters. Calculate the focal length of a telescope necessary for that the patient that allows the patient to read a text from a distance of 30 cm.

Answer: 5.45 cm.

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