НАУКАТА СРЕЩУ СТРАХА[1]

(The Science versus Fear)
Академик Тодор Николов
(Plenary report)
Глобални проблеми и глобални отговорности

"Изисква се смелост
да се страхуваш"
Монтен, "Опити", III, 6

"Не съм песимист. Според мен,
да виждаш злото там,
където то съществува,
е един вид оптимизъм"
Роберто Роселини

В забързаното ни всекидневие отвсякъде се носят предупреждения от най-различно естество, които могат да се съберат в думата "ВНИМАНИЕ!". При това тези апели се отнасят за най-различни опасности: от уличното движение и ширещият се бандитизъм до предстоящи природни катастрофи и/или до открития, от които ти настръхват косите. Вярно е, че новините за кризи, катастрофи, скандали и др.п. са повече не само поради мощното влияние на медиите (преса, телевизии, радио), а и поради факта, че те правят по-силно впечатление на обикновените хора в сравнение със съобщения за нормално протичащи събития и процеси[2].

Хората не признават, но те имат чувство за страх в света, в който живеят. Затова те леко възприемат всякакви идеи за предстоящи катастрофи, за апокалипсис сега, за края на света. Психолозите определят това като тип "катастрафално мислене". В тази връзка има един необясним парадокс: независимо от склонността към катастрофално мислене, в дългосрочен план хората не мислят за най-лошото. Очевидно и днес важи мисълта на Тит Ливий: "Краят на света няма да бъде утре".

Според мнозина политолози страхът е станал част от съвременните политически процеси. Динамиката на съвременното развитие и потокът от негативна информация, която залива обществото изискват от учените, отдадени на своите изследвания, да отделят подобаващо време за популяризиране на научните резултати, за борба срещу лъженауката и за осветляване на регионалните и глобалните проблеми, стоящи пред съвременността.

Съвременната епоха се отличава коренно от предходните етапи от историята на човечеството. Поне четири са главните отличителни черти на съвременността: ускореното развитие на науките и технологиите, наличието на

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[1] Основната част от този доклад е публикуван в Сп. на БАН, CXX, 1, 2007.

[2] Съществена е и ролята на журналистите. Ето един пример: в предаване на програма "Хоризонт" на БНР на 19.04.2000 г., 17,30 ч. водещата журналистка (Юлия Гигова) съобщава: "Днес преди обяд започна дискусия "Младите в науката", организирана от Съюза на учените, БАН и др. институции. Скандал нямаше, но това не значи, че проблемът е маловажен".

оръжия за масово поразяване, ужасяващото замърсяване на околната среда и демографският прираст. Всички тези особености на съвременността имат глобален характер.

Редица учени и религиозни дейци прибавят към тези фундаментални характеристики на нашата епоха нарушаване на основни етични принципи и изпъкващи противоречия и противоборства между етноси и религии, или това, което обобщено се нарича сблъсък на цивилизациите.

Във времето на глобализацията хората трябва да се обединят, защото имат обща съдба на Земята – съвместното съществуване или несъществуване. Глобалните проблеми изискват и глобални отговорности. И учените трябва да бъдат сред първите при формиране на новите хоризонти пред човечеството. Трябва да избегнем печалния и реален факт, че човек е първия биологически вид на Земята, който по израза на К. Сейгън "е разработил средства за своето унищожение". Безброй светли умове са мечтали за "Рай на Земята", но трябва да се осъзнае, че не може да има "Рай на Земята", ако няма "Мир на Земята".

Елементарната наука от древността, развивана от философите, е била свързана с търсене на обяснения за човека и околния свят. Тя се е развивала много бавно и без забележимо влияние върху общественото развитие.

Ще си позволя да цитирам кратките бележки на М. Борн за ранните корени на науката. "Атомната наука – пише Борн - започва около 600 години преди новата ера с разсъжденията на гръцките философи Талес, Анаксимандър и Анаксимен, които първи са се замислили над природата, водени от чисто любопитство и желание за знание, без непосредствена практическа полза. Атомистите Левкип и Демокрит постулират съществуването на природни закони и се опитват да обяснят разнообразието, проявявано от различните вещества чрез различно подреждане и движение на невидими малки, непроменими, и неделими частици — атомите.

Тази пленителна, красива и вдъхновяваща идея за същността на материята е била погребана в забрава за дълго време, защото е нямало средства за нейната проверка. Даже главната идея, че една теоретична конструкция може да се провери само чрез систематично експериментиране, е трябвало тепърва да се оформя и развива. Самите гърци са допринесли много за това. Ние им дължим не само основаването на абстрактната математика, но също така и първите ѝ приложения към физиката, например статиката на твърдото тяло и течностите, Птолемеевата система за небесните тела и др.

Гръцката цивилизация бе унищожена от нахлуване отвън. Но, арабите взимат и запазват научната традиция на гърците. Те я пренасят до народите в Европа, които стават водещи в науката от XVI столетие нататък. Разбира се, не бива да се забравя, че е имало периоди, в които науката е процъфтявала също така в Китай и Индия" (Борн, 1981, с. 54).

Науката «възкръсва от пепелта» преди около 300 години и оттогава се развива с ускорение. Тя бързо започва да влияе върху техниката и допринася за развитие на технологиите за да се стигне до съвременността, когато научните и технологични постижения са в основата на просперитета на много държави.

Днес науката е не само важна съставка на съвременната цивилизация, но тя е и сред най-важните ѝ характеристики. Това обаче не е основание за упреци към учените, че те, както казва един учен, са виновни за всичко – не само за атомната и водородната бомба, но и за лошото време.

Разбира се учените също носят своите отговорности. Нека припомним, че в ранната еволюция на човечеството интелектът е бил в основата на умения за създаване на оръжия и оръдия на труда. Този факт остава трайна характеристика на човека. В основата са интелектът и развиващите се на негова основа умения. В напредналия етап на развитие на науката за техническата реализация на научните открития спомага развитието на инженерната мисъл и практика. Така че учените и инженерите могат да се гордеят със своите открития, но и носят отговорност за своите разработки.

”Тоталната атака срещу глобалната околна среда – пише К. Сейгън - не трябва да се приписва единствено на жадните за печалба индустриалци или на тесногръдите и корумпирани политици.Самите ние също трябва да поемем голяма част от вината.

Племето на учените също е изиграло важна роля. Много от нас въобще не са си дали труда да се замислят за дългосрочните последствия от своите изобретения. Твърде лесно сме се съгласявали да връчим ужасяващи сили в ръцете на предложилия най-висока цена или на управниците на държавата, в която се е случило да живеем. В много случаи ни е липсвал моралният ориентир. Още от самото си начало философията и науката твърде много са искали — нека използвам думите на Рене Декарт - „да ни превърнат в господари и собственици на Природата". Или - както е казал Франсис Бейкън — искали сме да използваме науката, за да преклоним цялата природа „в услуга на Човека". Бейкън говори за „Човека", който упражнява своите „права над Природата". „Природата - пише Аристотел - е създала всички животни в името на човека." „Ако го нямаше човекът – заявява Имануел Кант, - цялото творение щеше да бъде просто пустош и щеше да е сторено напразно. “Съвсем доскоро слушахме за “овладяването” на природата и “превземането” на Космоса – сякъш природата и космосът са някакви врагове , които трябва да бъдат покорени.

Голяма е била и ролята на религиозното племе. Западните секти са проповядвали, че точно както ние трябва да се прекланяме пред Господ, така и цялата останала природа трябва да се подчинява на нас. А особено в съвременната епоха, изглежда, проявяваме много повече старание по отношение на втората половина на това предложение» отколкото се съобразяваме с първата. В реалния осезаем свят - което личи не от думите, а от делата ни - много хора явно се стремят да бъдат господари на Мирозданието - с някой епизодичен поклон, според изискванията на общественото приличие, пред някои модерен за момента бог. Както Декарт, така и Бейкън пишат под силното влияние на религията. Представата за „ние срещу Природата" ни е завещана от религиозната традиция. В книгата „Битие" Бог дава на човека ,,властта... над всяка жива твар” и ние трябва да внушаваме страх и ужас на „всеки звяр". Човекът е призован да „подчини" природата, като „подчини" е превод на една староеврейска дума със силен военен подтекст. В Библията има още много такива неща - както и в средновековната християнска традиция, от която се е зародила съвременната наука. В исляма, за разлика от това, не се забелязва тенденция природата да бъде обявявана за враг.

Разбира се, както науката, така и религията са сложни и многопластови структури, които обхващат много различни, дори противоречащи си мнения. Именно учените откриха глобалната екологична криза и привлякоха световното внимание върху нея, а има и такива, които - макар и на значителна лична цена - отказват за работят върху нещо, което може да навреди на техните събратя. От друга страна, именно в религията за първи път е формулиран императивът да бъде почитано всяко живо същество.

Вярно е, че в юдео-християно-мюсюлманската традиция няма нищо, което дори да се доближава до преклонението пред Природата в индуистко-будистко-джайниската традиция или това при автохтонното население на Америка. И наистина, както Западната религия, така и Западната наука са се отклонили от правия път и са стигнали до твърдението, че природата се явява само декор за нашата история, че е светотатство да обявим природата за свещенна” (Сейгън, 2005, с. 174)

Човек живее с убеждението, че е своеобразен “наместник” на Бога и е призван да се грижи за планетата. В тази връзка е уместно да припомним индианската мъдрост: ***“Не сме наследили Земята от своите предци, а сме я взели назаем от децата си”.***

Проблемът за отговорността на учените в съвременната епоха и специално по проблемите на различните заплахи за планетата ни се разглежда от много автори: Борн, 1981; Докинс, 1998; Браун, Флавин, Френч, 1999; Кенеди, 1995; Крайтън, 2005; Сейган, 2005; Тофлър, 2002; Хънтингтън, 2000; Clark, Carpenter, Barber et al., 2001 ; Palumbi, 2001 и мн. др. – списъкът е много дълъг.

Заедно с оценките и споделените тревоги повечето от авторите показват дълбок филисофски размисъл върху съвременното състояние на света и нашите отговорности. Бих искал да посоча като пример няколко автори.

Книгата на великия физик Макс Борн “Моят живот и възгледи” представлява проникновен разказ на един гениален учен, мъдър и мислещ човек за развитието на физиката в началото на миналия век и за неговото място като учен и университетски професор. Заедно с това тя е и дълбок анализ на принципите и моралните норми, които водят и трябва да водят учените. Според М. Борн “единственото, което може да ни спаси, е една стара мечта на човечеството: световен мир и световна организация”.

Книгата на британския учен-зоолог Ричард Докинс “Себичният ген” е забележителен анализ на генетичните дадености и поведенческите реакции на човека. Тя е апел “*да се учим* на благородство, защото ние сме родени егоисти”.

В последната си книга “Милиарди, милиарди”, завършена на предсмъртния му одър, големият американски учен – астроном Карл Сейгън разглежда актуалните проблеми на науката, на Земята и на Вселената. Отделните глави на книгата са написани под формата на есета и пораждат невероятни завладяващи мисли за природата около нас, за отговорността на учените, за перспективите на света. Заветът на К. Сейгън е: “да освободим разума си от догми... Нека проведем състезание по почтеност”.

На края на този повече или по-малко случаен списък ще спомена книгата на Майкъл Крайтън “Състояние на страх”. Крайтън е випускник на Харвардския университет, по специалност медицина; автор е на блестящи романи-бестселъри като “Щамът Андромеда”, “Жертвата”, “Изгряващото слънце”, “Джурасик парк” и др. Автор също на редица научни трудове в областта на биологията, медицината, информатиката и екологията. “Състояние на страх” е роман-художествена измислица, но пропит от много сериозни, макар и противоречиви размисли на героите и изключително ценни и точни бележки под линия, в които се цитират актуални научни статии от авторитетни списания. Основната нишка на повествованието на М. Крайтън в този роман е: защитата на природата. Талантът на Крайтън ни представя дискусиите по съвременните екологически проблеми под формата на трилър.

“Състояние на страх” е книга за лъженауката, която е маскирана под формата на глобален проблем. При това маскировката е толкова добра, че мнозина учени трудно преценяват доколко са обосновани основите на тази “нова теория”, която по инерция увлича голяма научна общност.

М. Крайтън не е учен, но той е забележителен белетрист, който следи научните достижения и се стреми да се ориентира в научните проблеми. Някои казват, че Крайтън и науката живеят в симбиоза. Затова не е учудващо, че американското списание “Wired” отбелязва следното: "дали ни харесва или не, но Крайтън просвещава американците, даже ако ние не четем неговите книги”.

Действително съвременните екологическите проблеми събират като във фокус усилията на учени, политици и общественици. Крайтън обръща внимание върху един много любопитен детайл, свързан с заплахите от замърсяването на околната среда. Случайно или не, но 1989 г. е показателна с белезите си на преход.

Въз основа на анализа на честотата на срещане на определени термини и схващания в новинарски емисии на основни телевизии в САЩ като Ен Би Си, Ей Би Си, Си Би Ес и на публукации във вестници в Ню Йорк, Вашингтон, Маями, Лос Анжелис и Сиатъл се вижда, че "през есента на 1989-а е настъпила сериозна промяна. Преди това медиите не са си падали по термини като криза, катастрофа, катаклизъм, чума или трагедия. Например през осемдесетте думата „криза" се е появявала в новинарските емисии приблизително толкова често, колкото и думата „бюджет". Освен това, преди 1989-а, прилагателни като краен, безпрецедентен, ужасяващ са били необичайни за телевизионните репортажи и вестникарските заглавия, Но след това всичко се променя". След 1989 г. "Тези термини стават все по-обичайни. Думата "катастрофа" се употребява пет пъти по-често през 1995-а, отколкото през 1985-а. А до двехиладната употребата й се удвоява още веднъж. Променят се и темите. Все по-голямо ударение се поставя върху страха, тревога
та, опасността, несигурността, паниката" (Крайтън, с. 438).

Не е пресилено да се каже, че днес светът се управлява със страх, което потвърждава думите на Ноам Чомски, че "много лесно е да манипулираш хората с помощта на страха".

Алюзията, която прави Крайтън в своя апел срещу страха е, че през есента на 1989 г. рухва Берлинската стена и това бележи края на Студената война. Няма значение дали формата, която използва Крайтън е чрез героите на своя роман, фактите бележат 1989 г. като преходна.

Нека припомним, че заплахите по време на Студената война бяха главно:

- Атомната бомба
- Водородната бомба
- Нови типове химическо и биологическо оръжие
- Спин
- Луда крава
- Генноинженерни технологии
- Африкански пчели-убийци
- Астероидни удари.

След 1989 г. заплахите главно са:

- Предстояща екологическа криза, която води до катастрофа
- Замърсяване на околната среда, масовото обезлесяване и особено изсичането на тропическите гори
- Увеличаване на CO_2
- Парников ефект
- Изтъняване на озонния екран
- Глобално затопляне
- Повишаване на океанското ниво
- Демографски взрив и пренаселеност на Земята
- Компютърен срив 2000
- Спин
- Птичи грип
- Пандемии

Крайъгълниат камък на тази серия от заплахи е **ГЛОБАЛНОТО ЗАТОПЛЯНЕ.** Реалността е такава, че всеки, който се противопостави на идеята за глобалното затопляне ще бъде анатемосан.

Интересно е, че първите аларми за глобално затопляне идват не от климатолози а от един философ и политолог – американецът Макс Каплан, който през 1974 г. обръща внимание към необходимостта от борба срещу глобалното затопляне. Едва през 1988 г. известният климатолог Джеймс Хансен акцентира върху опасността от този процес, искайки подкрепа от правителството на САЩ.

Идеята за глобалното затопляне бързо привлича множество привърженици – учени, политици, военни и става една от най-обсъжданите теми в последните десетина години, предмет както на международни научни форуми, на междуправителствени конференции и на форуми на ООН. Същевременно тази идея става основа и на най-чести спекулации и псевдонаучни прогнози за предстоящ апокалипсис, Всемирен потоп, поради глобално повишение на морското ниво и световна война за хранителни ресурси.

Привържениците на глобалното затопляне прогнозират, че най-късно до средата на XXI век глобалното морско ниво ще се повиши повече от 8 м.

Определено трябва да се каже, че подобни прогнози са абсурдни, тъй като климатът е изключително сложно нещо, което трудно се прогнозира защото зависи от множество фактори и се характеризира със сложни хаотично протичащи процеси и явления. Надеждни прогнози днес се правят за петдневен период, останалите са повече гадаене върху статистически вероятности. Във всеки случай никой не се опитва да прави прогнози за повече от десет дни напред. А някои, занимаващи се с компютърно моделиране предсказват каква ще бъде температурата след 100 или 1000 години. Разбира се, прогнози за далечни периоди са възможни и необходими, но те могат да очертаят само вероятностни тенденции в промяна на климата и то само на основата на задълбочени познания за фундаменталните фактори, които определят климата на нашата планета и особено на слънчевата активност.

Едно от най-мащабните и впечатляващи климатични явления на Земята е течението Ел Ниньо (ENSO-effect), което оказва колосално въздействие върху земния климат. То се проявява през интервал от около 9 години, но понякога възниква и приблизително на 4 години. Очевидно е, че компютърните климатични модели (поне засега) не могат да предскажат нито кога ще възникне, нито колко ще продължи Ел Ниньо. Затова дългосрочните прогнози на климата са безпочвени. А спекулациите с глобалното затопляне, както всички спекулации са осъдителни.

Непределеността по проблема за глобалното затопляне поражда силен скептицизъм и за грозящата опасност за живота върху Земята от бързото затопляне на климата.

Голямата опасност от глобалното затопляне се крие в потециала му за резки климатични промени. Енергетически то е по-изгодно от глобалното застудяване, но ще доведе по повишаване на глобалното морско ниво с действителни катастрофални последици за много острови, особено в Тихия и Индийския океан, а също и за повечето континентални крайбрежия.

Наложително е човечеството да намали до минимум своето отрицателно въздействие върху природната среда и да сътвори щадяща политика за защита на естествените екосистеми и атмосферата от замърсяване.

Още през 1827 г. френският физик Жозеф Фурие изказва мнението, че земната атмосфера изпълнява функция, подобна на стъклата на една оранжерия – въздухът пропуска слънчевата топлина, като не му позволява да се изпари обратно в Космоса. Парниковият ефект е естествено явление, без което не е възможно съществуването на живия свят. Ако не съществува този щит средната земна температура, която сега е 15^0 С била значително по-ниска – минус 19^0 С и животът на Земята практически ще замре.

Следва да се подчертае, че върху климата на Змеята определящо влияние оказват три основни орбитални величини, свързани с ротацията на Земята: *наклона* *на земната ос* (*обликвитета или осовата инклинация*), *прецесията* и *ексцентрицитета*. Тези астрономични промеливи величини, заедно с гравитационните въздействия на Слънцето и Луната върху Земята имат различна продължителност и ритмичност.

http://bg.wikipedia.org/wiki/%D0%97%D0%B5%D0%BC%D1%8F_%28%D0%BF%D0%BB%D0%B0%D0%BD%D0%B5%D1%82%D0%B0%29Наклонът на земната ос е 23,45°. Ориентацията на оста се запазва една и съща в различните части на земната орбита — ефект обуславящ земните сезони. Когато оста е наклонена в посока към Слънцето северното полукълбо получава повече светлина отколкото южното и там е лято, докато в обратния случай когато оста е наклонена в посока обратна на Слънцето южното полукълбо получава повече светлина. Прецесията на земната ос води до бавно изменение на настъпването на сезоните спрямо положението на Земята по нейната орбита — прецесия (изпреварване) на равноденствията. Ефекта на прецесията върху климата на планетата зависи от ексцентрицитета на земната орбита — колкото е по-голям то толкова по-резки изменения на климата могат да настъпят.

Много съществени за климата са т.нар. цикли от Слъчевата група . Наречени така по интензитета на слънчевата активност, която се изменя циклично през период от 11 години. Т.нар. магнитен цикъл на Слънцето е равен на два 11-гидишни цикъла. Съществуват и цикли с продължителност 50 и 90-100 години (т.нар. векови цикли); има и цикли с по-голяма продължителност 2200-2400 години. Слънчевата активност влияе върху атмосферата и магнитосферата и съответно върху живите организми. Възможно е в тази група да се отнася и явлението Ел Ниньо (ENSO - Ел Ниньо южни осцилации), което беше споменато по-горе. Неговите най-силни изяви са през интервал от около 10 г. (9.9 а).

Съществено влияние върху климата оказват и периодите, когато Слънчевата система пресича мощни прахови облаци, които намаляват слънчевата активност.

Интересна е астрономическата теория на сръбския геофизик и метеоролог Милутин Миланковичем, разработена през 20-те години на миналия век. Миланкович отделя три орбитални елемента, влияещи върху климата, отбелязани по-горе: 1) колебанието на земната ос, която (гледана отгоре) описва в пространството кръг приблизително през интервал от 25 Ка; 2) изменение на земната ос по отношение плоскостта на земната орбита (еклиптиката). Тези изменения стават с периодичност от 41 Ка; 3) третият елемент е свързан с изменението на формата на земната орбита от почти кръгова до леко овална (разтеглена) - елиптична.

Лекото глобално затопляне в края на XX в. е обусловено главно от астрономични (орбитални) фактори и не е повлияно съществено от човешката дейност, макар че ролята на човешкия фактор (индустриалната дейност и използването на изкопаемите горива) нараства със застрашаваща скорост.

По данни на британски учени (цитирани от BBC – 14.12.2000 г.) средната глобална температура на земната повърхност в последните три десетилетия е нараствала средно с $0,2^0$ C на всеки десет години, но в периода от 1945 до 1975 г. включително не е имало затопляне. Повишение на температурата се наблюдава между 1910 и 1945 г., след което до средата на 70-те години температурата е устойчива. Според същият колектив британски учени затоплянето ще продължи и през следващите няколко десетки години. Те отбелязват също, че над 80 % от наблюдаваните колебания на глобалната средна температура, а също над 60 % от колебанията на земната температура за период от 10 до 50 години са причинени от външни (астрономични и орбитални фактори).

По най-нови данни през 2005 г. средната температура на земната повърхност е нараснала с 0,42 % и се е повишила до 14^0 С. Тази тенденция продължава вече 23 години. Според българския астроном Борис Комитов топлият климат на Земята след 1975 г. е свързан с най-високо ниво на слънчевата активност през последните 800 години.

Независимо от тези факти не може да се обеснява всичко с глобалното затопляне: от смърта на делфина във водите на Темза до промяна на наклона на земната ос. Но нали наклонът на земната ос е сред фундаменталните фактори, определящи климата върху Земята. Този наклон обяснява и това, че на Земята съществува различие на климата в северното и южното полукълбо. Освен това, макар че бързото изменение на климата сега се нарича "глобално затопляне", то включва в себе си и глобално захлаждане (илюстрация -johnstonsarchive.net). Мнението на някои автори, че глобалното затопляне буквално променя въртенето на Земята и намалява продължителността на деня е също безпочвено, тъй като тези феномени са с фундаментални характеристики и не могат да се повлияят от кратковременни явления и процеси. Още повече промяната в скоростта на въртенето на Земята, намаляването на броя на дните в годината и на тяхната продължителност е известно още от палеозойската ера (500 Ма ВР) и е обяснено от законите на Кеплер за небесната механика.

Не можем да не се съгласим с М. Крайтън, който казва, че идеята за глобалното затопляне не е просто една научна хипотеза или теория. Самата тя създава криза и е призив за действие. "А една криза трябва да се изследва, проучванията трябва да се финансират, нужни са политически и бюрократични структури по целия свят. И за нула време огромен брой метеоролози, геолози, океанографи изведнъж се преквалифицираха в „климатолози", заети с управлението на въпросната криза".

Независимо от всичко друго, около идеята за глобалното затопляне се въртят много пари. Известно е, че само смяната на фреоните носи грамадни доходи на крупните химически и хладилни компании. Има данни, че смяната на хладилниците в само в САЩ е донесло на производителите около 220 млрд. долара.

ГЛОБАЛНО ЗАТОПЛЯНЕ ИЛИ ГЛОБАЛНО ЗАСТУДЯВАНЕ?

Съществува и мнение, че идеята за глобалното затопляне е необоснована хипотеза. На нея се противопоставя хипотезата за глобалното застудяване. Двата процеса обаче са в диалектическа връзка.

По принцип затоплянето на океаните трябва да доведе до затоплянето на сушата, но връзката е много по-сложна. Студените течения, които са по-плътни поради солеността си, се спускат на дълбочина, отправяйки се към тропичните ширини, където изтласкват топлата вода на повърхността. Топлите води се връщат към Арктика. Гълфстрим, включен в тези процеси, подгрява климата на Британските острови средно с 5^0 С. Общото повишаване на температурата на океана, свързано с глобалното затопляне довежда до това, че конвенционалният обмен на теченията (студено-топло) губи своята основна характеристика и студените течения вече не са в сдъстояние да подтикват и да дават сила и направление на топлите течения. По такъв начин Гълфстрим е с намален обем и по-слаб (поради по-малкия приток на студени води от север), което постепенно довежда до понижаване на средногодишната температура на Британските острови до 11^0 С, а в някои региони на островите температурите стават като в норвежкия архипелаг Сволборг (на 1200 км от Северния полюс).

Учени от различни страни отбелязват, че предстои период на намалена слънчева активност, която ще причини глобално захлаждане на климата на Земята. Такава идея развива нашия учен Б. Комитов (2001), който прогнозира глобално захлаждане приз 21 век.

В началото на тази година РИА Новости (6.02.2006 г.) разпространи съобщение, че според руския астроном Хабибулло Абдусаматов – сътрудник на Пулковската астрономическа обсерватория, в средата на 21 век се очаква глобално понижаване на температурите. Това всъщност е тезата и на българския астроном Б. Комитов.

Х. Абдусаматов отбелязва резултатите от анализа на дълговременните изменения на 11-годишните вариации на слънчевата активност, които показват, че глобалното затопляне на Земята е достигнало своя максимум. По-нататък ще последва намаляване на слънчевата активност и бавно постепенно намаляване на глобалната температура на Земята. Така глобалното затопляне ще се последва от глобално застудяване. Началото на понижаването на глобалната температура се очаква да бъде към 2012-2013 г. Към 2035-2045 г. слънчевата светимост ще достигне своя минимум, след който с известна закъснение (инерция) от 15-20 години ще настане поредния климатичен минимум – значително глобално застудяване. Нормален ход на нормален (естествен) процес.

Има безспорни данни, че <u>в глобален мащаб се наблюдава леко повишаване на средната годишна температура на Земята</u>. Тази естествена тенденция се утежнява от силното замърсяване на околната среда и атмосферата, свързано с дейността на човека. Ако не се ограничи отрицателното въздействие на човека върху природата можем със сигурност да кажем, че човечеството ще понесе катастрофални последици с неизброими човешки жертви и огромни материални щети. Това няма да бъде края на света, но ще остане дълбока следа в геоложкия летопис, която ще бъде съизмерима с епохата, през която измират динозаврите.

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Какви са главните изводи от проблемите, свързани с основни природни процеси, в т.ч. и опасни тенденции в развитието на климата, на състоянието на околната среда, на използването на природните ресурси и др. п.? Има ли основания да се плашат хората?

При всички случаи е необходимо по-голямо международно съгласие по глобалните проблеми, стоящи пред човечеството днес. И управление на процесите с оглед да се осигури не само икономически просперитет, но и да се опази равновесието с околната среда.Защото, както отбелязва българския учен П. Пенчев (2001) "<u>лошото управление е многократно по-лошо от природни бедствия</u>".

Дявол знае защо съвременното общество живее при постоянно насаждане на различни страхове, а науката не е достатъчно силна за да противодейства на непрестанната лавина от предупреждения за все нови и нови заплахи. Не бива да викаш "Пожар" в претъпкан театър, а защо може да крещиш непрекъснато "СПИН", "птичи грип", "предстоящ апокалипсис" от страниците на пресата и електронните медии?

Няма никакво основание да се плашат хората, но е необходимо по-широко и балансирано образование, което ще осигури познания света около нас. Една от бедите днес е, че повечето хора са почти неграмотни за основните закони, които регулират равновесието в природата.

Освен това трябва да се има предвид, че основната причина за заплахите срещу природата се бедността и алчността, а те трудно ще бъдат изкоренени. Те формират много сложни, преплетени взаимоотношения между севера и юга, между изтока и запада, които често са като в джунгла, а както казва М. Борн (1981) от джунглата не може да възникне нова етика.

За преодоляване на основните трудности, свързани със състоянието на природната среда са необходими съвременни познания както в хуманитарните, така и по природните науки. Само така могат да се преодолеят заблуждения и предразсъдъци, защото обикновено хората, по израза на Монтен, вярват на това, което познават най-малко.

Кризите и катастрофите са част от еволюцията на Земята и на населяващия организмов свят, за която е характерна определена цикличност. Глобалните цикли обикновено са с по-голяма продължителност (100 Ка и повече). Така че, можем да приемем, че на Земята има белези на криза във връзка със замърсяването на околната среда и с глобалното затопляне, но няма никакви основания на тръбим, че предстоят близки катастрофални промени на нашата планета.

Ще завърша с мисълта на Ричард Файнман **"Науката е усвоеното от човечеството умение да не се заблуждава"**. Поради това ролята на учените в съвременния свят е да развиват и популяризират знанията, а с това ще се изправят срещу неоправданите страхове, насаждани от различни "информатори" и "популяризатори" чрез прогнози за предстоящ апокалипсис.

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NEW NONEXISTENCE RESULTS FOR SPHERICAL 5-DESIGNS

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Abstract. We investigate the structure of spherical 5-designs of relatively small cardinalities. We obtain some bounds on the extreme inner products of such designs. As a result, in 42 cases we prove nonexistence of designs of corresponding parameters. Our approach can be applied for other strengths and cardinalities.

1 Introduction

The spherical designs were introduced in 1977 by Delsarte-Goethals-Seidel [8].

Definition 1. *A spherical τ-design $C \subset \mathbb{S}^{n-1}$ is a finite nonempty subset of $\mathbb{S}^{n-1}$ such that*

$$\frac{1}{\mu(\mathbb{S}^{n-1})}\int_{\mathbb{S}^{n-1}} f(x)d\mu(x) = \frac{1}{|C|}\sum_{x\in C} f(x) \qquad (1)$$

($\mu(x)$ is the Lebesgue measure) holds for all polynomials $f(x) = f(x_1, x_2, \ldots, x_n)$ of degree at most τ (i.e. the average of $f(x)$ over the set C is equal to the average of $f(x)$ over $\mathbb{S}^{n-1}$). The number $\tau = \tau(C)$ is called strength of C.

Denote by $B(n,\tau)$ the minimum possible cardinality of a τ-design on $\mathbf{S}^{n-1}$, i.e.

$$B(n,\tau) = \min\{|C| : C \in \mathbf{S}^{n-1} \text{ is a } \tau\text{-design}\}.$$

Delsarte-Goethals-Seidel [8] prove the following lower bound for $B(n,\tau)$, i.e.

$$B(n,\tau) \geq D(n,\tau) = \begin{cases} 2\binom{n+e-2}{n-1}, & \text{if } \tau = 2e-1, \\ \binom{n+e-1}{n-1} + \binom{n+e-2}{n-1}, & \text{if } \tau = 2e. \end{cases}$$

In this paper we prove nonexistence of certain spherical 5-designs. This does not give direct improvement to the above bound for $\tau = 5$ but sheds some light on the problem for existence of designs of prescribed dimension, strength and cardinality.

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The following equivalent definition of spherical designs is very suitable for our purposes.

Definition 2. *A spherical* τ*-design* $C \subset \mathbb{S}^{n-1}$ *is a finite nonempty subset of* $\mathbb{S}^{n-1}$ *such that for any point* $x \in C$ *and any real polynomial* $f(t)$ *of degree at most* τ*, the equality*

$$\sum_{y\in C\setminus\{x\}} f(\langle x,y\rangle) = f_0|C| - f(1) \qquad (2)$$

holds, where f_0 *is the first coefficient in the expansion of* $f(t) = \sum_{i=0}^{k} f_i P_i^{(n)}(t)$ *in terms of the Gegenbauer polynomials [1].*

We are interested in the following:

Problem. Given dimension n and cardinality M decide whether a 5-design on $\mathbb{S}^{n-1}$ of cardinality $|C| = M$ exists.

We obtain restrictions on the structure of 5-designs of relatively small cardinalities, i.e. close to $D(n,5) = n(n+1)$. This allows us to obtain some nonexistence results. Our method can be applied for other odd strengths and cardinalities.

All known constructions of spherical designs (see, for example, [2, 3, 9, 11]) suggest that the structure of the design with respect to any of its points should be investigated. This can be done by using suitable polynomials in (1) combined with some geometric arguments.

In Section 2 we describe our approach. The results are formulated in general but will be used for $\tau = 5$. In fact, we continue investigations started in [6, 5, 4] with proving nonexistence of designs in many cases. The results for $\tau = 5$ in dimensions $n \le 25$ are presented in sections 3 and 4.

It was proved in [5] that the condition $\rho_0|C| \ge 2$ is necessary for the existence of τ-designs $C \subset \mathbb{S}^{n-1}$ with odd τ and $|C|$. For 5-designs, we prove (ruling out 42 cases) that in dimensions $5 \le n \le 25$ this can be replaced by $\rho_0|C| > 3$.

2 Preliminaries

Let $C \in \mathbb{S}^{n-1}$ be a spherical τ-design, where $\tau = 2e-1 \ge 3$ is odd. For every point $x \in C$ we consider the inner products of x with all other points of C, i.e.

$$I(x) = \{\langle u,x\rangle : u \in C \setminus \{x\}\} = \{t_1(x), t_2(x), \ldots, t_{|C|}(x)\},$$

where $-1 \le t_1(x) \le t_2(x) \le \cdots \le t_{|C|-1}(x) < 1$. Using suitable polynomials in (1) we obtain lower and upper bounds for the extreme inner products in $I(x)$ for some special points x. Let us recall some results from [10, 5, 4].

It follows from [10] (see also [5]) that for every fixed cardinality $|C| \geq D(n, 2e-1)$ there exist uniquely determined real numbers $-1 \leq \alpha_0 < \alpha_1 < \cdots < \alpha_{e-1} < 1$ and $\rho_0, \rho_1, \ldots, \rho_{e-1}$, $\rho_i > 0$ for $i = 0, 1, \ldots, e-1$, such that the equality

$$f_0 = \frac{f(1)}{|C|} + \sum_{i=0}^{e-1} \rho_i f(\alpha_i) \tag{3}$$

is true for every real polynomial $f(t)$ of degree at most $2e-1$. We use (3) in some calculations of $f_0|C| - f(1)$ in the right hand side of (2). Another useful formula for f_0 is

$$f_0 = a_0 + \sum_{i=1}^{\lfloor k/2 \rfloor} \frac{a_{2i}(2i-1)!!}{n(n+2)\cdots(n+2i-2)} = a_0 + \frac{a_2}{n} + \frac{3a_4}{n(n+2)} + \cdots, \tag{4}$$

where $f(t) = a_0 + a_1 t + a_2 t^2 + \cdots + a_k t^k$.

The numbers α_i, $i = 0, 1, \ldots, e-1$ are all roots of the equation

$$P_e(t)P_{e-1}(s) - P_e(s)P_{e-1}(t) = 0,$$

where $P_i(t) = P_i^{(n-1)/2,(n-3)/2}(t)$ is a Jacobi polynomial [1]. The weights ρ_i can be calculated by

$$\rho_i = -\frac{\prod_{0 \leq j \leq e-1, j \neq i}(1-\alpha_j^2)}{\alpha_i |C| \prod_{0 \leq j \leq e-1, j \neq i}(\alpha_i^2 - \alpha_j^2)}.$$

Theorem 1. [5] *If $C \subset \mathbb{S}^{n-1}$ is a τ-design with odd $\tau = 2e-1$ and odd $|C|$ then $\rho_0|C| \geq 2$.*

Lemma 1. [5] *Let $C \subset \mathbb{S}^{n-1}$ be a τ-design with odd $\tau = 2e-1$. For any point $x \in C$ we have $t_1(x) \leq \alpha_0$ and $t_{|C|-1}(x) \geq \alpha_{e-1}$. If $|C|$ is odd then there exist a point $x \in C$ such that $t_2(x) \leq \alpha_0$.*

Lemma 2. [4] *Let $C \subset \mathbb{S}^{n-1}$ be a τ-design with odd $\tau = 2e-1$ and of odd cardinality $|C|$. Then there exist three distinct points $x, y, z \in C$ such that $t_1(x) = t_1(y)$ and $t_2(x) = t_1(z)$. Moreover, we have $t_{|C|-1}(z) \geq \max\{\alpha_{e-1}, 2\alpha_0^2 - 1\}$.*

It is convenient to use the following notation: $U_{\tau,i}(x)$ (respectively $L_{\tau,i}(x)$) for any upper (resp. lower) bound on the inner product $t_i(x)$. When a bound does not depend on x we omit x in the notation. For example, the first bound from Lemma 1 is $t_1(x) \le U_{\tau,1} = \alpha_0$ and the last bound from Lemma 2 is $t_{|C|-1}(z) \ge L_{\tau,|C|-1}(z) = \max\{\alpha_{e-1}, 2\alpha_0^2 - 1\}$.

3 General bounds

In what follows we take $\tau = 5$. Let $C \subset \mathbb{S}^{n-1}$ be a 5-design of odd cardinality $|C|$ and $x, y, z \in C$ be points as in Lemma 2. Then α_0, α_1 and α_2 are the roots of the equation

$$n(1-\alpha)((n+2)(n+3)\alpha^2 + 4(n+2)\alpha - n + 1) = 2\alpha|C|(3 - (n+2)\alpha^2).$$

We denote $g(t) = (t-\alpha_1)^2(t-\alpha_2)^2$.

After [4], there are 42 open cases where $5 \le n \le 25$, $|C|$ is odd and $2 \le \rho_0|C| \le 3$. In all these cases $2\alpha_0^2 - 1 > \alpha_2$ and Lemma 2 gives

$$t_{|C|-1}(z) \ge L_{5,|C|-1}(z) = 2\alpha_0^2 - 1. \qquad (5)$$

We focus on the inner products in $I(x)$ and $I(z)$. The main purpose is obtaining a upper bound

$$t_1(z) \le U_{5,1}(z) < \alpha_0. \qquad (6)$$

We start with a lower bound on $t_1(z) = t_2(x)$.

Lemma 3. *We have* $t_1(z) \ge L_{5,1}(z)$ *where* $L_{5,1}(z)$ *is the smallest root of the equation* $2g(t) = \rho_0|C|g(\alpha_0)$.

Proof. Use (2) with $g(t)$ for x and C and (3) for the left hand side. For $g(t)$ we have $g(t_i(x)) \ge 0$ for $i \ge 3$, $g(t_1(x)) \ge g(t_2(x))$, because $g(t)$ is decreasing in $(-\infty, \alpha_1)$ and $t_1(x) \le t_2(x) \le \alpha_0$. Therefore

$$\begin{aligned}\rho_0 g(\alpha_0)|C| &= g_0|C| - g(1) = \sum_{i=1}^{|C|-1} g(t_i(x)) \\ &\ge g(t_1(x)) + g(t_2(x)) \ge 2g(t_2(x)) = 2g(t_1(z)),\end{aligned}$$

since $t_1(z) = t_2(x)$ by Lemma 2. Now the conclusion follows by using again that $g(t)$ is decreasing in $(-\infty, \alpha_1)$.. $\square$

We illustrate our method with two examples which appear in parts after each important assertion. The numerical results are approximated as follows: the lower bounds are rounded up and the upper bounds are truncated as we usually give six digits after the decimal point. We do the same in all numerical applications. The calculations were performed by a MAPLE programme. Both the programme and the calculations for every separate case can be obtained from the authors upon request.

In all appearances of Example 1 (resp. Example 2) we treat the case $n=11$ and $|C|=147$ (resp. the case $n=17$ and $|C|=343$). Both examples contain the complete nonexistence proofs for the corresponding designs.

Example 1. For $n=11$ and $|C|=147$, we have $\alpha_0=-0.830399$, $\alpha_1=-0.248366$ and $\alpha_2=0.293051$. Then the equation from Lemma 3 is approximated by $(t+0.248366)^2(t-0.293051)^2=0.582577$ whose smallest root is approximately $-0.892289.$ Therefore $t_1(z)\geq L_{5,1}(z)=-0.892289$. $\diamond$

Example 2. Analogously, for $n=17$ and $|C|=343$, we approximate $\alpha_0=-0.816081$, $\alpha_1=-0.210608$, $\alpha_2=0.240641$ and $t_1(z)\geq L_{5,1}(z)=-0.892617$. $\diamond$

Lemma 3 allows us to obtain a good upper bound on $t_2(z)$.

Lemma 4. *We have* $t_2(z)\leq U_{5,2}(z)$ *with* $U_{5,2}(z)$ *defined in the proof.*

Proof. We denote $q(t)=t^2+at+b$ and use (2) with $f(t)=(t-t_2(z))q^2(t)$ for z and C where the parameters a and b will be determined later but have to be such that the polynomial $q(t)$ has two real roots in $[\alpha_0,\alpha_2]$. We have $f(t_i(z))\geq 0$ for $i\geq 2$, $f(t_1(z))\geq f(L_{5,1}(z))$ because $f(t)$ is increasing in $(-\infty,t_2(z))$. Therefore

$$\begin{aligned} f_0|C|-f(1) &= \sum_{i=1}^{|C|-1} f(t_i(z)) \\ &\geq f(t_1(z))+f(t_{|C|-1}(z))\geq f(L_{5,1}(z))+f(L_{5,|C|-1}(z)) \end{aligned}$$

(we use (5) and $t_1(z)\geq L_{5,1}(z)$ by Lemma 3). This gives the following inequality for $t_2(z)$

$$t_2(z)\leq F(a,b)=\frac{A(a,b)}{B(a,b)},$$

where

$$A(a,b)=\frac{6a|C|}{n(n+2)}+\frac{2ab|C|}{n}-q^2(1)-L_{5,1}(z)q^2(L_{5,1}(z))-L_{5,|C|-1}(z)q^2(L_{5,|C|-1}(z))$$

and

$$B(a,b)=\frac{3|C|}{n(n+2)}+\frac{(a^2+b)|C|}{n}+b^2|C|-q^2(1)-q^2(L_{5,1}(z))-q^2(L_{5,|C|-1}(z)).$$

After the optimization over a and b we obtain the bound $t_2(z) \le U_{5,2}(z)$. □

Example 1. (Continued) We have $t_2(z) \le U_{5,2}(z) = -0.774411$. ◇

Example 2. (Continued) We have $t_2(z) \le U_{5,2}(z) = -0.744010$. ◇

Lemma 5. *We have*

$$t_{|C|-1}(x) \ge L_{5,|C|-1}(x) = U_{5,1}(z)U_{5,2}(z) - \sqrt{(1 - U_{5,1}^2(z))(1 - U_{5,2}^2(z))}, \quad (7)$$

where $U_{5,i}(z)$, $i = 1, 2$, *is any (good) upper bound for* $t_i(z)$, $i = 1, 2$.

Proof. Denote by φ and ψ the acute angles such that $\cos\varphi = -U_{5,1}(z)$ and $\cos\psi = -U_{5,2}(z)$. Let $u \in C$ be such that $\langle u, z\rangle = t_2(z)$. Then the angle between the vectors x and u is at most $\varphi + \psi$ and we have

$$\begin{aligned} t_{|C|-1}(x) &\ge \langle x, u\rangle \ge \cos(\varphi + \psi) \\ &= U_{5,1}(z)U_{5,2}(z) - \sqrt{(1 - U_{5,1}^2(z))(1 - U_{5,2}^2(z))}. \end{aligned} \qquad □$$

Lemma 6. *We have* $t_3(z) \ge \min\{L_{5,3}(z), \alpha_1\}$, *where* $L_{5,3}(z)$ *is the smallest root of the equation* $2g(t) = (\rho_0|C| - 1)g(\alpha_0) - g(L_{5,|C|-1}(z))$.

Proof. Use (2) with $g(t)$ for z and C. Applying similar arguments as in Lemma 1 and assuming $t_3(z) < \alpha_1$, we consecutively obtain

$$\begin{aligned} \rho_0 g(\alpha_0)|C| &= g_0|C| - g(1) = \sum_{i=1}^{|C|-1} g(t_i(z)) \\ &\ge g(t_1(z)) + g(t_2(z)) + g(t_3(z)) + g(t_{|C|-1}(z)) \\ &\ge g(\alpha_0) + 2g(t_3(z)) + g(L_{5,|C|-1}(z)). \end{aligned}$$

This implies the assertion since $g(t)$ is decreasing in $(-\infty, \alpha_1)$. □

Example 1. (Continued) We have $t_3(z) \ge L_{5,3}(z) = -0.801894$. ◇

Example 2. (Continued) We have $t_3(z) \ge L_{5,3}(z) = -0.807530$. ◇

In all 42 cases under consideration we have the following ordering for our parameters and the bounds from Lemmas 1-6:

$$-1 < L_{5,1}(z) < \alpha_0 < L_{5,3}(z) < U_{5,2}(z) < \alpha' < \alpha_1 < \alpha_2 < L_{5,|C|-1}(z),$$

where α' is the smallest root of the derivative $f'(t)$ of the polynomial $f(t) = (t - \alpha_0)g(t)$.

We further consider several cases for the location of the inner products $t_2(z)$ and $t_3(z)$. The details will be shown in the next section.

4 The location of $t_2(z)$ and $t_3(z)$

Using the bounds from Lemmas 1-6 we consider two cases for the location of $t_2(z)$ with respect to α_0.

4.1 Case 1: $t_2(z) \in [\alpha_0, U_{5,2}(z)]$

We are ready to obtain better upper bound on $t_1(z)$ as required by (6).

Lemma 7. *We have $t_1(z) \leq U_{5,1}(z)$, where $U_{5,1}(z)$ is the smallest root of the equation $f(t) = (\rho_0|C| - 1)f(\alpha_0) - f(L_{5,|C|-1}(z))$ and $f(t) = (t - L_{5,3}(z))g(t)$.*

Proof. Use (2) with $f(t)$ for z and C. We have

$$\begin{aligned}\rho_0|C|f(\alpha_0) &= f_0|C| - f(1) = \sum_{i=1}^{|C|-1} f(t_i(z)) \\ &\geq f(t_1(z)) + f(t_2(z)) + f(t_{|C|-1}(z)) \\ &\geq f(t_1(z)) + f(\alpha_0) + f(L_{5,|C|-1}(z)),\end{aligned}$$

which implies the inequality $t_1(z) \leq U_{5,1}(z)$ since $f(t)$ is increasing in $(-\infty, \alpha_0)$. □

Remark. According to (6) we need $U_{5,1}(z) < \alpha_0$. This is the case when

$$\rho_0|C| < 2 + \frac{f(2\alpha_0^2 - 1)}{f(\alpha_0)}.$$

Example 1. (Continued) We have $t_1(z) \leq U_{5,1}(z) = -0.852885$. ◇

Example 2. (Continued) We have $t_1(z) \leq U_{5,1}(z) = -0.829252$. ◇

Having a good upper bound $t_1(z) \leq U_{5,1}(z)$ we are in a position to obtain strong necessary condition for the existence of C. We use $t_2(x) = t_1(z) \leq U_{5,1}(z)$ and $t_{|C|-1}(x) \geq L_{5,|C|-1}(x)$ by Lemmas 2 and 5 respectively.

The next Lemma gives a necessary condition for the existence of C. It uses the information about $I(x)$ which is collected so far. We denote shortly this check for existence by check (x).

Lemma 8. (Check for existence by x) *If there exist* $a, b \in [\alpha_0, \alpha_2]$ *such that*

$$\text{check}(x) := h_0|C| - h(1) - 2h(U_{5,1}(z)) - h(L_{5,|C|-1}(x)) < 0,$$

where $h(t) = (t-a)^2(t-b)^2$, *then* C *does not exist.*

Proof. Use (2) with $h(t)$, x and C. We have

$$\begin{aligned} h_0|C| - h(1) &= \sum_{i=1}^{|C|-1} h(t_i(x)) \geq h(t_1(x)) + h(t_2(x)) + h(t_{|C|-1}(x)) \\ &\geq 2h(t_2(x)) + h(L_{5,|C|-1}(x)) \geq 2h(U_{5,1}(z)) + h(L_{5,|C|-1}(x)) \end{aligned}$$

(we use $t_1(x) \leq t_2(x) = t_1(z) \leq U_{5,1}(z)$ and $t_{|C|-1}(x) \geq L_{5,|C|-1}(x)$), which implies the assertion since $h(t)$ is decreasing in $(-\infty, \alpha_0)$ and increasing in $(\alpha_2, +\infty)$. □

Example 3. For $n = 12$, $|C| = 171$ we have $t_1(z) \leq U_{5,1}(z) = -0.887772$, $L_{5,|C|-1}(x) = 0.501028$ and $\text{check}(x) = -0.038759 < 0$. Therefore 5-designs on $\mathbb{S}^{11}$ with 171 points such that $t_2(z) \in [\alpha_0, U_{5,2}(z)]$ do not exist. In fact this case was ruled out by Boumova-Boyvalenkov-Danev in [4]. ◇

After the optimization of a and b we can still have $\text{check}(x) \geq 0$ (converse to the inequality from Lemma 8). Then we continue with a recursive procedure which replaces α_0 with $U_{5,1}(z)$ whenever possible in Lemma 7 and again turn to Lemma 8 with better $U_{5,1}(z)$ and $L_{5,|C|-1}(x)$.

Example 1. (Continued) At the first step we have $t_1(z) \leq U_{5,1}(z) = -0.852885$, $L_{5,|C|-1}(x) = 0.330162$ and $\text{check}(x) = 0.195233 > 0.$ The second step gives $t_1(z) \leq U_{5,1}(z) = -0.872394$, $L_{5,|C|-1}(x) = 0.366334$ and $\text{check}(x) = 0.099869 > 0$ again, but the third step gives $t_1(z) \leq U_{5,1}(z) = -0.903327$, $L_{5,|C|-1}(x) = 0.428156$ and $\text{check}(x) = -0.072318 < 0$. Therefore Lemma 8 implies that 5-designs on $\mathbb{S}^{10}$ with 147 points such that $t_2(z) \in [\alpha_0, U_{5,2}(z)]$ do not exist. ◇

Example 2. (Continued) Similarly, after six steps we obtain $t_1(z) \le U_{5,1}(z) = -0.926518$, $L_{5,|C|-1}(x) = 0.437940$ and $\mathrm{check}(x) = -0.218084 < 0$. Therefore 5-designs on $\mathbb{S}^{16}$ with 343 points such that $t_2(z) \in [\alpha_0, U_{5,2}(z)]$ do not exist. ◇

This approach rules out 40 cases out of the 42 under consideration. The two remaining cases ($n = 7$, $|C| = 63$ and $n = 8$, $|C| = 81$) are ruled out by precise consideration how close is $t_3(z)$ to $L_{5,3}(z)$.

We have $t_2(z) \in [\alpha_0, U_{5,2}(z)]$ and consider two possibilities for $t_3(z)$.

Case 1.1. Let us have $t_3(z) \in [L_{5,3}(z), L_{5,3}(z) + \varepsilon]$, where $\varepsilon > 0$ is a positive number such that $L_{5,3}(z) + \varepsilon < \alpha_1$. We have the analog of Lemma 8 (check for existence which uses the information about $I(z)$).

Lemma $8'$. (Check for existence by z) *If there exist $a, b \in [\alpha_1, \alpha_2]$ such that*

$$\mathrm{check}(z) := h_0|C| - h(1) - h(U_{5,1}(z)) - 2h(L_{5,3}(z) + \varepsilon) - h(L_{5,|C|-1}(z)) < 0,$$

where $h(t) = (t-a)^2(t-b)^2$, then C does not exist.

Proof. Use (2) with $h(t)$, z and C. We have

$$\begin{aligned} h_0|C| - h(1) &= \sum_{i=1}^{|C|-1} h(t_i(z)) \ge h(t_1(z)) + h(t_2(z)) + h(t_3(z)) + h(t_{|C|-1}(z)) \\ &\ge h(t_1(z)) + 2h(t_3(z)) + h(L_{5,|C|-1}(z)) \\ &\ge h(U_{5,1}(z)) + 2h(L_{5,3}(z) + \varepsilon) + h(L_{5,|C|-1}(z)) \end{aligned}$$

(we use $t_1(z)U_{5,1}(z)$, $t_2(z) \le t_3(z) \le L_{5,3}(z) + \varepsilon$ and $t_{|C|-1}(z) \ge L_{5,|C|-1}(z)$), which implies the assertion since $h(t)$ is decreasing in $(-\infty, \alpha_1)$ and increasing in $(\alpha_2, +\infty)$. □

After the optimization of a and b we can still have $\mathrm{check}(z) \ge 0$ (converse to the inequality from Lemma $8'$). Then we continue with a recursive procedure which replaces α_0 with $U_{5,1}(z)$ whenever possible in Lemma 7 and again turn to Lemma $8'$ with better $U_{5,1}(z)$ and $L_{5,|C|-1}(z)$. With $\varepsilon = 0.008$ this approach rules out the two remaining cases $n = 7$, $|C| = 63$ and $n = 8$, $|C| = 81$. *Case 1.2.* Let us have $t_3(z) \ge L_{5,3}(z) + \varepsilon$, where $\varepsilon = 0.008$ as above. We have the analog of Lemma 7 for obtaining to a better upper bound on $t_1(z)$.

Lemma $7'$. *We have* $t_1(z) \leq U_{5,1}(z)$, *where* $U_{5,1}(z)$ *is the smallest root of the equation* $f(t) = (\rho_0|C| - 1)f(\alpha_0) - f(L_{5,|C|-1}(z))$ *and* $f(t) = (t - L_{5,3}(z) - \varepsilon)g(t)$.

Proof. Use (2) with $f(t)$ for z and C. We have

$$\begin{aligned}\rho_0|C|f(\alpha_0) &= f_0|C| - f(1) = \sum_{i=1}^{|C|-1} f(t_i(z)) \\ &\geq f(t_1(z)) + f(t_2(z)) + f(t_{|C|-1}(z)) \\ &\geq f(t_1(z)) + f(\alpha_0) + f(L_{5,|C|-1}(z)),\end{aligned}$$

which implies the assertion since $f(t)$ is increasing in $(-\infty, \alpha_0)$. □

We check for existence by Lemma 8 for the point x. After finding the optimal values of a and b we can still have the converse inequality, i.e. $\mathrm{check}(x) \geq 0$. Then we continue with a recursive procedure which replaces α_0 with $U_{5,1}(z)$ whenever possible in Lemma $7'$ and again turn to Lemma 8 with better $U_{5,1}(z)$ and $L_{5,|C|-1}(x)$. This rules out the two remaining cases $n = 7$, $|C| = 63$ and $n = 8$, $|C| = 81$.

Thus we finally have obtained the nonexistence of all 42 designs under consideration assuming $t_2(z) \in [\alpha_0, U_{5,2}(z)]$.

4.2 Case 2: $t_2(z) \in [t_1(z), \alpha_0]$

We have $t_2(z) \in [t_1(z), \alpha_0] \subseteq [L_{5,1}(z), U_{5,2}(z) := \alpha_0]$. We can not obtain good bounds $t_1(z) \leq U_{5,1}(z)$ at this point. This is why we start with investigation of the location of $t_3(z)$ with respect to α'.

Case 2.1. Let us have $t_3(z) \in [L_{5,3}(z), \alpha']$. We start with new lower bounds on $t_2(z)$ and $t_3(z)$.

Lemma 9. *We have* $t_2(z) \geq L_{5,2}(z)$, *where* $L_{5,2}(z)$ *is the smallest root of the equation* $2g(t) = \rho_0|C|g(\alpha_0) - g(\alpha') - g(L_{5,|C|-1}(z))$.

Proof. Use (2) with $g(t)$ for z and C. We have

$$\begin{aligned}\rho_0 g(\alpha_0)|C| &= g_0|C| - g(1) = \sum_{i=1}^{|C|-1} g(t_i(z)) \\ &\geq g(t_1(z)) + g(t_2(z)) + g(t_3(z)) + g(t_{|C|-1}(z)) \\ &\geq 2g(t_2(z)) + g(\alpha') + g(L_{5,|C|-1}(z)),\end{aligned}$$

which implies the assertion since $g(t)$ is decreasing in $(-\infty, \alpha_1)$. □

Lemma 10. *We have* $t_3(z) \geq \min\{L_{5,3}(z), \alpha_1\}$, *where* $L_{5,3}(z)$ *is the smallest root of the equation* $g(t) = (\rho_0|C| - 2)g(\alpha_0) - g(L_{5,|C|-1}(z))$.

Proof. Using $t_2(z) \leq \alpha_0 = U_{5,2}(z)$ as in Lemma 6 we have

$$\begin{aligned}\rho_0 g(\alpha_0)|C| &= g_0|C| - g(1) = \sum_{i=1}^{|C|-1} g(t_i(z)) \\ &\geq g(t_1(z)) + g(t_2(z)) + g(t_3(z)) + g(t_{|C|-1}(z)) \\ &\geq 2g(\alpha_0) + g(t_3(z)) + g(L_{5,|C|-1}(z)),\end{aligned}$$

which implies the assertion since $g(t)$ is decreasing in $(-\infty, \alpha_1)$. □

Remark. A new better bound $t_1(z) \geq L_{5,1}(z)$ can be obtained but we have not found its applications.

In all cases we have $L_{5,2}(z) \leq t_2(z) \leq \alpha_0 \leq L_{5,3}(z) \leq t_3(z) \leq \alpha'$ which seems to be a strong restriction.

Example 1. (Continued) We have $\alpha' = -0.680699$. Then Lemmas 9-10 give $t_2(z) \geq L_{5,2}(z) = -0.858038$ and $t_3(z) \geq L_{5,3}(z) = -0.769776$.
◇

Example 2. (Continued) Analogously, we have $\alpha' = -0.664851$ and $t_2(z) \geq L_{5,2}(z) = -0.860278$ and $t_3(z) \geq L_{5,3}(z) = -0.798687$.
◇

Now, we are in a position to obtain a upper bound $t_1(z) \leq U_{5,1}(z)$ as required by (6).

Lemma 11. *We have* $t_1(z) \leq U_{5,1}(z)$, *where* $U_{5,1}(z)$ *is the smallest root of the equation* $f(t) = -f(L_{5,2}(z)) - f(L_{5,3}(z)) - f(L_{5,|C|-1}(z))$, *where* $f(t) = (t - \alpha_0)g(t)$.

Proof. Use (2) with $f(t)$ for z and C. We have

$$0 = f_0|C| - f(1) - \sum_{i=1}^{|C|-1} f(t_i(z)) \ge f(t_1(z)) + f(t_2(z)) + f(t_3(z)) + f(t_{|C|-1}(z))$$
$$\ge f(U_{5,1}(z)) + f(L_{5,2}(z)) + f(L_{5,3}(z)) + f(L_{5,|C|-1}(z)),$$

which implies the assertion since $f(t)$ is increasing in $(-\infty, \alpha_0)$. We note the inequality $f(t_3(z)) \ge f(L_{5,3}(z))$ which follows by $t_3(z) \in [L_{5,3}(z), \alpha']$ and explains our choice to work with α'. □

Lemma 12. *If there exist* $a, b \in [\alpha', \alpha_2]$ *such that* $\mathrm{check}(z) := h_0|C| - h(1) - h(U_{5,1}(z)) - h(\alpha_0) - h(\alpha') - h(L_{5,|C|-1}(z)) < 0$, *where* $h(t) = (t-a)^2(t-b)^2$, *then* C *does not exist.*

Proof. Use (2) with $h(t)$, z and C. We have

$$h_0|C| - h(1) = \sum_{i=1}^{|C|-1} h(t_i(z)) \ge h(t_1(z)) + h(t_2(z)) + h(t_3(z)) + h(t_{|C|-1}(z))$$
$$> h(U_{5,1}(z)) - h(\alpha_0) - h(\alpha') - h(L_{5,|C|-1}(z)),$$

which implies the assertion. □

As in Case 1 we apply a recursive procedure. We come back consecutively to Lemmas 9-11 and check (z) for existence by Lemma 12, while $\mathrm{check}(z) \ge 0$.

Example 1. (Continued) We have $\alpha' = -0.680699$. The first step gives $t_2(z) \ge L_{5,2}(z) = -0.858038$, $t_3(z) \ge L_{5,3}(z) = -0.769776$, $t_1(z) \le U_{5,1}(z) = -0.848575$ and $\mathrm{check}(z) = 0.079302 > 0$ but the second step gives $t_2(z) \ge L_{5,2}(z) = -0.856549$, $t_3(z) \ge L_{5,3}(z) = -0.739912$, $t_1(z) \le U_{5,1}(z) = -0.874519$ and $\mathrm{check}(z) = -0.009398 < 0$. Therefore 5-designs on $\mathbb{S}^{10}$ with 147 points such that $t_2(z) \in [t_1(z), \alpha_0]$ and $t_3(z) \in [L_{5,3}(z), \alpha']$ do not exist. ◇

Example 2. (Continued) We have $\alpha' = -0.664851$. The first step gives $t_2(z) \ge L_{5,2}(z) = -0.860278$, $t_3(z) \ge L_{5,3}(z) = -0.798687$ and $t_1(z) \le U_{5,1}(z) = -0.775811$. This is one of the bad cases when $U_{5,1}(z) > \alpha_0$. ◇

The last procedure rules out 18 out of all 42 cases. The remaining 24 cases (including Example 2) are resolved by a precise consideration how close is $t_3(z)$ to $L_{5,3}(z)$ as in the end of Case 1. More precisely we have the following two subcases.

Case 2.1.1. Let us have $t_3(z) \in [L_{5,3}(z), L_{5,3}(z)+\varepsilon]$, where $\varepsilon > 0$ is such that $L_{5,3}(z)+\varepsilon < \alpha'$. Now we have new upper bound $t_3(z) \leq U_{5,3}(z) = L_{5,3}(z)+\varepsilon$ and analogs of Lemma 11 and Lemma 12.

Lemma 11′. *We have* $t_1(z) \leq U_{5,1}(z)$, *where* $U_{5,1}(z)$ *is the smallest root of the equation* $f(t) = -f(L_{5,2}(z)) - f(L_{5,3}(z)) - f(L_{5,|C|-1}(z))$, *where* $f(t) = (t-\alpha_0)g(t)$.

Proof. Use (2) with $f(t)$ for z and C. We have

$$
\begin{aligned}
0 = f_0|C| - f(1) = \sum_{i=1}^{|C|-1} f(t_i(z)) &\geq f(t_1(z)) + f(t_2(z)) + f(t_3(z)) + f(t_{|C|-1}(z)) \\
&\geq f(t_1(z)) + f(L_{5,2}(z)) + f(L_{5,3}(z)) + f(L_{5,|C|-1}(z)),
\end{aligned}
$$

which implies the assertion since $f(t)$ is increasing in $(-\infty, \alpha_0)$. □

Lemma 12′. *If there exist* $a, b \in [\alpha', \alpha_2]$ *such that* $\mathrm{check}(z) := h_0|C| - h(1) - h(U_{5,1}(z)) - h(\alpha_0) - h(U_{5,3}(z)) - h(L_{5,|C|-1}(z)) < 0$, *where* $h(t) = (t-a)^2(t-b)^2$, *then* C *does not exist.*

Proof. Use (2) with $h(t)$, z and C. We have

$$
\begin{aligned}
h_0|C| - h(1) &= \sum_{i=1}^{|C|-1} h(t_i(z)) \geq h(t_1(z)) + h(t_2(z)) + h(t_3(z)) + h(t_{|C|-1}(z)) \\
&\geq h(U_{5,1}(z)) + h(\alpha_0) + h(U_{5,3}(z)) + h(L_{5,|C|-1}(z)),
\end{aligned}
$$

which implies the assertion. □

A recursive procedure (as described above) using Lemmas 9-10, $11', 12'$ rules out the remaining 24 cases when $t_2(z) \in [t_1(z), \alpha_0]$ and $t_3(z) \in [L_{5,3}(z), L_{5,3}(z)+\varepsilon]$. For 22 of them $\varepsilon = 0.01$ works. For the remaining two cases ($n = 19$, $|C| = 427$) and ($n = 21$, $|C| = 519$), we need $\varepsilon = 0.008$ and $\varepsilon = 0.007$ respectively.

Example 2. (Continued) We have $\alpha' = -0.664851$ and $\varepsilon = 0.01$. The first step gives $t_2(z) \geq L_{5,2}(z) = -0.860278$, $t_3(z) \geq L_{5,3}(z) = -0.798687$, $t_3(z) \leq U_{5,3}(z) = L_{5,3}(z) + \varepsilon = -0.788687$, $t_1(z) \leq U_{5,1}(z) = -0.826848$ and $\mathrm{check}(z) = -0.006815 < 0.$ Therefore 5-designs on $\mathbb{S}^{16}$ with 343 points such that $t_2(z) \in [t_1(z), \alpha_0]$ and $t_3(z) \in [L_{5,3}(z), L_{5,3}(z)+0.01]$ do not exist. ◇

Case 2.1.2. Let us have $t_3(z) \in [L_{5,3}(z) + \varepsilon, \alpha']$, where ε is as above. Then we apply the analog of Lemma 11 and check(z) by Lemma 12.

Lemma 11 $''$. *We have* $t_1(z) \le U_{5,1}(z)$, *where* $U_{5,1}(z)$ *is the smallest root of the equation* $f(t) = -f(L_{5,2}(z)) - f(L_{5,3}(z) + \varepsilon) - f(L_{5,|C|-1}(z))$, *where* $f(t) = (t - \alpha_0)g(t)$.

Proof. Use (2) with $f(t)$ for z and C. We have

$$0 = f_0|C| - f(1) - \sum_{i=1}^{|C|-1} f(t_i(z)) \ge f(t_1(z)) + f(t_2(z)) + f(t_3(z)) + f(t_{|C|-1}(z))$$
$$\ge f(t_1(z)) + f(L_{5,2}(z)) + f(L_{5,3}(z) + \varepsilon) + f(L_{5,|C|-1}(z)),$$

which implies the assertion since $f(t)$ is increasing in $(-\infty, \alpha_0)$. □

A recursive procedure with Lemmas 9-10, $11''$ and 12 resolves all remaining 24 cases when $t_2(z) \in [t_1(z), \alpha_0]$ and $t_3(z) \in [L_{5,3}(z) + \varepsilon, \alpha']$.

Example 2. (Continued) Again, we have $\alpha' = -0.664851$ and $\varepsilon = 0.01$. After six steps we obtain $t_2(z) \ge L_{5,2}(z) = -0.852473$, $t_3(z) \ge L_{5,3}(z) = -0.694411$, $t_1(z) \le U_{5,1}(z) = -0.885379$ and $\mathrm{check}(z) = -0.028429 < 0$. Therefore 5-designs on $\mathbb{S}^{16}$ with 343 points such that $t_2(z) \in [t_1(z), \alpha_0]$ and $t_3(z) \in [L_{5,3}(z) + 0.01, \alpha']$ do not exist. ◇

This ends Case 2.1 with a nonexistence proof for all 42 designs under consideration under the assumption $t_2(z) \in [t_1(z), \alpha_0]$ and $t_3(z) \in [L_{5,3}(z), \alpha']$.

Case 2.2. Let us have $t_3(z) > \alpha'$, i.e. $t_3(z) \ge L_{5,3}(z) = \alpha'$. We obtain immediately a upper bound $t_1(z) \le U_{5,1}(z)$ as required by (6).

Lemma 13. *We have* $t_1(z) \le U_{5,1}(z)$, *where* $U_{5,1}(z)$ *is the smallest root of the equation* $2f(t) = \rho_0|C|f(\alpha_0) - f(L_{5,|C|-1}(z))$, *where* $f(t) = (t - \alpha')g(t)$.

Proof. Use (2) with $f(t)$ for z and C. We have

$$\rho_0|C|f(\alpha_0) = f_0|C| - f(1) - \sum_{i=1}^{|C|-1} f(t_i(z))$$
$$\ge f(t_1(z)) + f(t_2(z)) + f(t_{|C|-1}(z)) \ge 2f(t_1(z)) + f(L_{5,|C|-1}(z)),$$

which implies the assertion. □

We now apply check(x) by using the better bound

$$L_{5,|C|-1}(x) = U_{5,1}(z)\alpha_0 - \sqrt{(1-U_{5,1}^2(z))(1-\alpha_0^2)}$$

in Lemma 8 using $t_2(z) \le \alpha_0 = U_{5,2}(z)$.

If $\mathrm{check}(x) \ge 0$ we continue with a recursive procedure which replaces α_0 with $U_{5,1}(z)$ whenever possible and again turn to check(x) with better $U_{5,1}(z)$ and $L_{5,|C|-1}(x)$.

Example 1. (Continued) We need eight steps to obtain $t_3(z) \ge \alpha' = -0.680699$, $t_1(z) \le U_{5,1}(z) = -0.890144$, $L_{5,|C|-1}(x) = 0.401037$ and $\mathrm{check}(x) = -0.011722 < 0$. This completes the proof in the last case. Therefore there exist no 5-designs on $\mathbb{S}^{10}$ with 147 points. ◇

Example 2. (Continued) Similarly, we have $\alpha' = -0.664851$ and $\varepsilon = 0.01$. We need twelve steps to obtain $t_1(z) \le U_{5,1}(z) = -0.890753$, $L_{5,|C|-1}(x) = 0.359055$ and $\mathrm{check}(x) = -0.015316 < 0$. This completes the proof in the last case. Therefore there exist no 5-designs on $\mathbb{S}^{16}$ with 343 points. ◇

This procedure rules out 36 out of all 42 cases. The last 6 cases are now ruled out by a precise consideration how close is $t_3(z)$ to α' and how close is $t_2(z)$ to α_0). We omit the details.

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ON WEB BASED TESTS AND ONLINE SURVEY

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***Abstract**: This article presents a Web System for Psychological Re-search (WSPR) developed for Department of Psychology (South-West University) as a faculty project. The system is capable of handling psycho-logical tests specifics (there are not a correct/wrong answers, factorization etc.). Commonly used e-Learning systems (Moodle, Claroline etc.) do not support such elements. The business logic of test preparation in WSPR is based on relaying question/answer sum of values with the results ranges.*

INTRODUCTION

This article discusses techniques for preparation and execution of web based tests and online questionnaires. It analyzes the implementation and integration of such a system in an university web space.

Major attention is focused on the system named WSPR (Web System for Psychological Research).

WSPR is created as South-West University interdisciplinary project in early 2005 on classical Open Source platform: Linux operating system, Apache web server and MySQL database server. The programming language is PHP.

A lot is said about advantages and disadvantages of this platform. Mainly discussion is about operating system independence (PHP is ported for Linux, Windows and some UNIX systems), PHP as object-oriented language, level of support, backward compatibility. The most important is that the platform is free – so this is the right choice for low cost projects [1].

The system aims to optimize the process of psycho-diagnostics. It is necessary as part of improvement lifecycle of scientific, research and educational process in the University.

We must note that the development of high quality web-based tests is expensive. So we must consider what our primary goal is: excellence or cost efficiency [2]. In our project there were financial limitations. They exclude excellence as primary project objective.

The details of the psycho-diagnostic process are not subject of this paper and will be addressed only when it is necessary for the explanation of the WSPR system specifics.

By using software for psycho-diagnostics, many difficulties, related to the conventional testing methods can be avoided. Such as: expenses for hard copy materials, slow and unsafe result processing, special measures for document preservation, difficult and many times impossible juxtaposition of results. Some of the advantages are obvious: Instant feedback; flexible time and location; great reliability in comparison with human made evaluation.

These difficulties are solved by the introduced web based system for psychological research.

The main tasks are:

To be assured access of researchers (teachers) and the objects of the research (students) with username and password.

The Mechanisms of test preparation must satisfy the following conditions:

• Factorization is required (factorization is when in one test different types of questions are presented. The answers of these questions provide information of different aspects of the object's personality – aggressiveness, sincerity, etc. – called factors).

• The questions are in a survey like format – here is no right or wrong answer.

• The opportunity of using multiple-choice single answers and multiple-choice multiple answers in the same test.

• The researcher can manipulate student's access to the test. The test is active (visible) or not.

• The question can be a picture or other image.

• The answer can be an image too.

• When the system evaluates answers of different test or factor, the weights of different answers are summarized.

• Every question is associated with no more than one factor, consequently the sum of the weights of the answers, given by the subject of research (the student), determines strictly the result of the corresponding factor.

Additional requirement is tests history. When a student makes a test twice or more the researcher can use previous results to observe the advancement.

Other feature of the product is multilanguage support and localization. The basic version of the product supports two languages: Bulgarian and English. The current language in use is defined in configuration file.

The system is written in a way, which provides convenient interface for internationalization and localization, thus facilitating the integration and support of new languages in the future.

Database Structure

The database relational schema consists of eight tables.

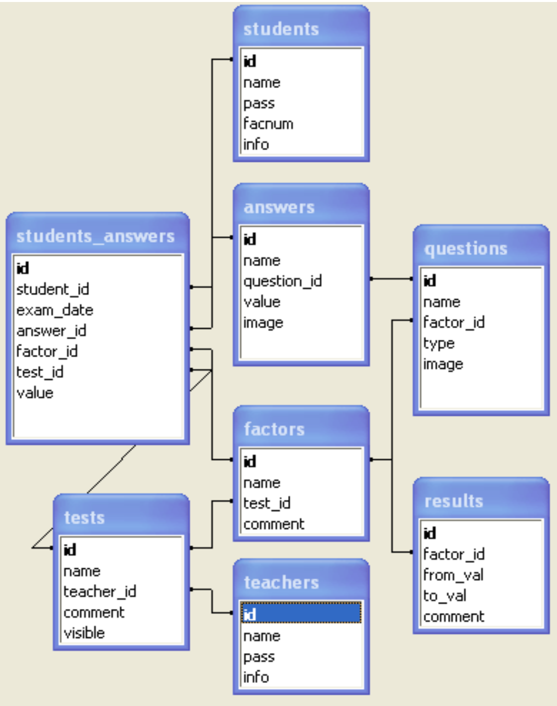

Fig. 1: The database relational schema

Two tables are used for storing the user information. The first one - **“teachers”** possess information about the researcher – in common case the teacher. There are two fields containing the username and the password of the teacher and one field for addition information about the teacher.

The second table - **“students”** is for second type of users – subjects of the research – students. It differs from the **“teachers”** table with one additional field – ‘facnum’, holding the faculty number of the student.

Once the teacher has a valid account he can create a new test. Every test has name, comment, status (visible or invisible) and owner. The ownership can be tracked down by [tests].[teacher_id] field which is in relationship with [teachers].[id].

The next table is **“factors”**. It is obvious that [factors].[test_id] is in relationship with [tests].[id]. Factors have name and comment fields. If the test has no factors, one is automatically assigned. That is because the results correlate with factors instead of the tests. Other fields in **“results”** table are ‘from_val’ and ‘to_val’ which define the range of values. For this range the ‘comment’ is assigned.

The **“questions”** table has five fields. [questions].[id] is the primary key. Every table has such field. The question is in [questions].[name] (for example “Do you believe in God?”). The field [questions].[type] has two possible values (0, 1) for two different types of questions – multiple-choice single answer questions and multiple-choice multiple answer questions. The **“questions”** table is in relationship with **“factors”** because questions are assigned to different factors (not to tests). The next table is **“answers”**. It is in relationship with **“questions”**. Both have the field [image] where the name of uploaded image is preserved. So the questions and the answers can be graphical objects. Every answer has a value – integer number.

For the same factors must be valid:

A – all answers
A1 – given answers
ai – answer’s value
R – all results

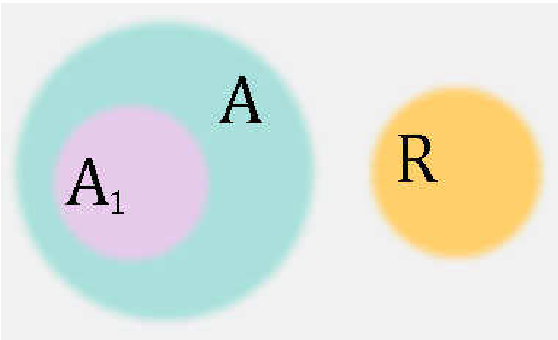

Fig. 2: Answers and Results Sets.

$$A_1 \subseteq A$$

$$a_1 \ldots a_n \in A_1$$

$$r_1 \ldots r_m \in R$$

$$a_1 + a_2 + \ldots + a_n \leq r_{\max}$$

$$a_1 + a_2 + \ldots + a_n \geq r_{\min}$$

When a student takes a test, his answers are written in **“students_answers”** table. There is a [test_id] field which is step back from database normalization because we can have the test identification number through the relationship [tests].[id] -> [factors].[test_id] and then [factors][id] -> [students_answers].[factor_id]. This time normalization doesn’t lead to excellent performance [3]. Other methods are twice and more slower than the simple selection (table 1).

Table 1. UNION; LEFT, RIGHT and INNER JOIN productivity.

SELECT * from students_answers

QUERY	Exec. time
, factors WHERE …	0,0019 sec.
LEFT JOIN factors ON(…	0,0019 sec.
RIGHT JOIN factors ON(..	0,0025 sec.
INNER JOIN factors ON (…	0,0018 sec.
WHERE test_id=…	0,0009 sec.

The described way for increasing the performance is important because when student’s answers are written we can generate reports based on these answers.

Business Logic

No specific definition is given in computers terminology for Business Logic but it is often used to note the tier that contains the application server in multitiers architecture of web application [4].

Here are three user roles in our application: administrator, researcher (most often this is the teacher) and the subject of the research (usually the students).

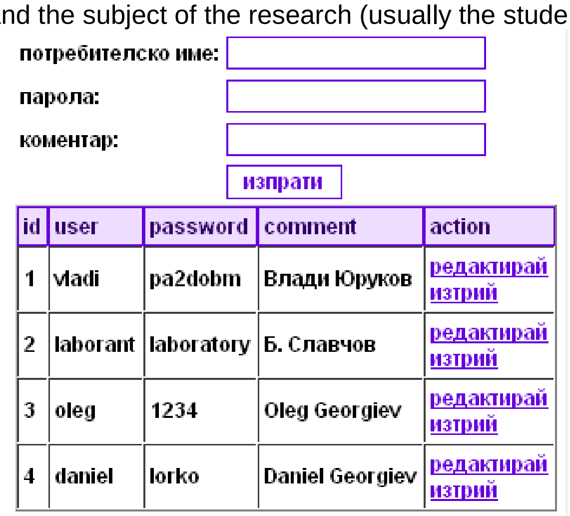

Fig. 3: Researcher’s login administration.

Administrator's user role is separated from the other part of the web application due to security reasons.

The main part of administrator's duties is the management of the username and password Access Control Lists (ACL) for students and teachers (Picture 2). The ACL management interface for students is similar to the management of teacher's ACL.

The Web Administrator can manage the system using the configuration file "conf/conf.php".

Table 2. conf/conf.php file

row	code
1	<?php
2	$lang = "bulgarian";
3	$db = "wspr";
4	$dbhost = "localhost";
5	$dbuser = "wspr";
6	$dbpass = "wspr@123";
7	$uploaddir = "./wspr/pics/";
8	$StartPage = "http://sharper.swu.bg/wspr/";
9	$dbfailed_message = "Connection failed!";
10	?>

The $lang variable contains the name of the language directory. All "lang*.php" files are in this directory, so adding support for new language can be done by translating all files to the new language for example in French and put them in "French" directory. At the time of this writing "Bulgarian" and "English" directories are present. In a specific moment only one locale can be used.

Rows 3, 4, 5 and 6 (Table 2) hold the database connection information.

Row 7 has the uploaddir: where the pictures have to be uploaded. This directory must have write permission for the system user used to run the apache web server.

These settings should be changed when moving the system to another server , the uploaddir should be writable and the connection to the database updated.

Other Web Administrator's duties are the regular system software and hardware support and backups.

Researcher's job role includes test creation. Basic actions with **tests** are adding, editing and removing. Other actions that the researcher can perform are changing the visibility status and reviewing the results. The answers given by the subject of the research can be seen in that section.

The next level is about **factors**. Again we have add, edit and delete. According to the relational model there are links to the management of **questions** and results. Only basic actions (add, edit and delete) for **answers** are present.

In answers and questions when an image is required it is uploaded with specific name. The name for images about questions includes the UNIX time stamp and the leading factor_id. The name of the answer's image uses the question_id and again the UNIX time stamp. Thus we can avoid concurrent file names.

Student's job role is not actually for students only. We are calling student the subject of the research, because in South-West University the system is used in department of Psychology but this doesn't mean that only students can be analyzed with WSPR.

Once the student logs in to the system, he sees a list with visible tests. The list consists of the test name, short instruction (or description) and the researcher's name. If he chooses to run a test, he must process it at once. All his answers are written into the **"students_answers"** table.

The system successfully runs "yes/no" tests, graphical tests, "a-never, b-sometimes, c-often, d-always" tests etc. Actually the system can run every test if enough results are provided because we can recognize every answer combination, given by the subjects of the research.

Generally there are three types of tests: single choice answers, multiple choice answers and mixed. If in a test there are n questions and they are q1, q2, …, qn for every question there are various number of answers (ai):

1) $f_{count}(q_i) = a_i,\ i=1, 2, \ldots, n$

For single choice answers tests, all combinations of given answers would be:

2) $a_1.a_2\ldots a_n$

For multiple choice answers tests all combinations of given answers would be:

3) $2^{a_1}.2^{a_2}\ldots 2^{a_n}$

So we can assume that for the mixed type of tests all combinations would be:

4) $a_1.a_2\ldots a_i.2^{(a_{i+1}+\ldots+a_n)}$

Where ***i*** is the number of single choice answer questions and ***[i+1, n]*** are multiple choice answer questions. As we can see 3) and 2) are special cases of 4). In 2) ***i = n*** and in 3) ***i = 0***.

Table 3. Test example with answer values

q_1		q_2	
Answer	value	answer	value
a)	1	a)	4
b)	2	b)	8
		c)	16

To achieve maximum diversity in answer combinations and to assure the uniqueness of every combination we are offering answers to be valued with powers of two (Table 3). If q1 and q2 are single choice answer questions all combinations are:

$a_1 = f_{count}(q_1) = 2$

$a_2 = f_{count}(q_2) = 3$

$a_1\ a_2 = 6$

and for all combinations, matching result values are:

sum(1, 4) = 5; sum(2, 4) = 6;
sum(1, 8) = 9, sum(2, 8) = 10;
sum(1, 16) = 17; sum(2, 16) = 18;

In that case powers of two is a "wasting" approach. But let face the multiple choice answer questions:

a1 = fcount (q1) = 2
a2 = fcount (q2) = 3
2(a1+ a2) = 25 = 32

And for all combinations, matching result values are:
sum() = 0; sum(1) = 1;
sum(1, 4) = 5; sum(1, 5) = 6;
sum(1, 2, 4, 8, 16) = 31;

We have proved that we can calculate answer combinations and for every combination we can construct unique weight (sum of values) by using powers of two.

The described system gives to the researcher the ability to unite different answer combinations under common result if the combinations are logically close enough ([results].[from_va] and [results).[to_val] form a range).

CONCLUSIONS

We examined one low cost CMS used for specific work with flexible structure and capable to cover wide variety of areas. When the “low cost” criterion precedes the “excellence” it affects the system with some restrictions and weaknesses. To note strengths and weaknesses we are using SWOT analysis approach.

Strengths:

1. Low cost of the project (total cost of the project was 2500 BGN. 900 BGN of them for computing only).
2. Open Source projects are used for the entire project lifecycle.
3. The system is flexible and covers wide variety of tests.
4. The system differs from the rest of e-Learning systems – they all had a right and wrong answer.
5. We use job roles. That makes a step forward for better security and management.
6. We had an interested customer before putting the system in production (department of Psychology, South-West University).
7. Hosting and administration after the projects end are provided (SWU).

Weaknesses:

1. The primary criterion is “low cost” not “excellence”.
2. It is in researchers hands to put a test in production.
3. The system had not been planned under e-Learning standards.
4. The students proceed all questions in a test at once instead of every question separately. That leads two different weaknesses: The system does not read the execution time for every question; linear tests are

Opportunities:

1. End-users respond to new technologies.
2. The testing process is optimized. It is a part of educational lifecycle
3. Always can be developed and added new tests.
4. There is no restriction in the number of the subjects of the research.

Threads:

1. The system is vulnerable to major competition (open source or commercial projects).
2. Technical support and administration can be discontinued from the supporting organization.

3. Security risks can always be mentioned when we deal with web based software application.

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DEFENSE MECHANISMS AGAINST COMPUTER ATTACKS "DISTRIBUTED DENIAL OF SERVICE" TYPE

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***Abstract**: According to statistical information, published by SANS institute (System Administration, Networking and Security) [1], one of the most critical and devastating classes of computer attacks is that of the type (Distributed Denial of Service, (DDoS), aimed at interfering in the accessibility to information resources. These attacks are accomplished by the combined actions of variety of program components available on Internet hosts. One of the current tasks in the field of computer safety is the development of relevant security methods against DDoS attacks and creating of well-grounded recommendations to choose from for the most effective means in the particular conditions. The article contains research of the existing and perspective mechanisms of defense against DDoS attacks.*

***Key words:** Distributed Denial of Service, Network Security.*

1. INTRODUCTION

Typical examples of the Denial of Service attacks are the well-known attacks WinNuke or SYN-Flood. They are characterized by sending of incorrectly formed web packet or by sending by evil-minded people of a large number of special packets, whose processing takes on all resources of domain's controller, which blocks the processing of the other requests.

What is the underlying principle of the SYN-Flood attack? A virtual connection is established so that two hosts can connect in the net, according to TCP protocol. When there is a request for connection by one of the hosts (for example the customer), a TCP – packet with a set flag SYN is sent to the other host (for example server). The server responds to the request with set flags SYN and ACK and a virtual connection is made after a confirmation by the customer has been received. In this case, if the confirmation by the customer has not arrived, the server is waiting for the response for a definite time, using for this purpose part of its resources. The attack SYN-Flood consists in sending of a large number of SYN-packets to establish a connection without the corresponding confirmation. As a result, the host engages too much resource for a non-existing connection and cannot process other requests, i.e., the efficiency of the host is disturbed.

Special systems or inter-network filters to protect the components of computer systems can be used for detecting of attacks, type Denial of Service. The system Real Secure of ISS Company can be named as an example of a means of detecting attacks. It is installed in the host under the control of Windows NT or Solaris and not only does it detect all attacks on the information system, it also prevents their influence on the operation of the elements of the computer system. The protective walls (Firewalls) operate in a similar way but they can repulse far less attacks.

DDoS attacks differ from DoS in principle. When they appear, the concept of attacking changes and the existing means of detecting of IDS (Intrusion Detection System) prove to be ineffective in most of the cases. The attacks are already accomplished at three levels. It can be seen from Fig. 1 that the evil-minded person (hacker) does not interact directly with the victim. He/She acts with the help of Masters and Daemons. There are usually several Masters installed in the heavy traffic hosts (with a broad channel in the Internet). They remain unnoticed when the transmitted/received information increases. Another important moment is that they are often severs (always switched on) and a restart of the computer is needed to clear the Master. The detecting of the Daemons follows after one or several Masters have been identified. A Daemon is a program of the Troy Horse type, which is installed on an alien machine (often using the well-known faults in the software). Having been installed on the machines, the Daemons get in touch with one of the Masters and receive commands from it. In this case the attack can be stopped by the neutralization of the Daemons. It is a problem that the computer with the Daemon in it does not suspect anything. The only possible way out of this situation is to handle the control over the Master, but in the more modern versions of software it is used for the realization of distributed attacks. The Daemon may not obey the Master's commands after the start of the attacks. The latter increases the probability of successful realization of the attack. But there is a drawback for the hacker because he or she will not be able to use this net of daemons for a further attack. While it is enough to add only a rule for the filtration of the inter-network screen under the DoS attack, the distributed attacks need several thousand rules (the attack is carried out by several thousand Daemons simultaneously). Most inter-network screens are not able to process such quantity of rules.

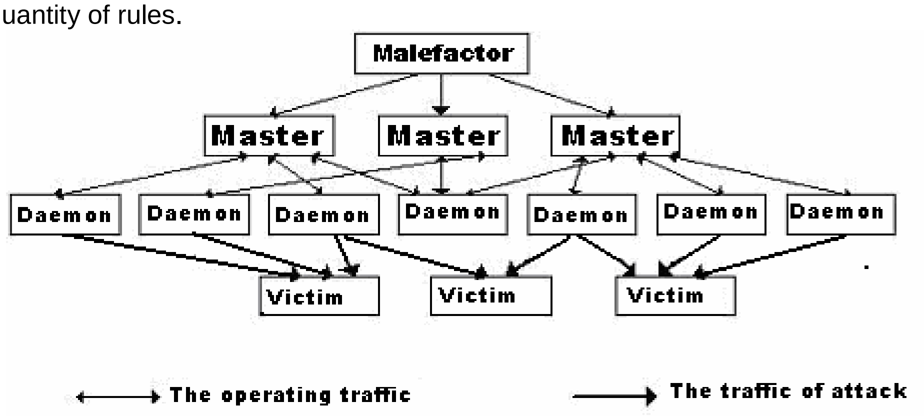

Fig. 1 DDoS attacks

The general approach to defense against DDoS attacks includes the realization of the following elements:

1) Defense against DDoS attacks;
2) Detection of the attack;
3) Identifying of the source of attacks;
4) Counteraction to the attack.

The defense includes setting of inter-network screens, testing if the system to make clear the weak points, taking measures to improve the server's protection, such as ban of all kinds of traffic uncritical to the server, optimization of the net infrastructure, etc.

Detecting of the fact of the attack on the basis of identification of anomalies in the operation of the server or detecting of misuse, as well as determining of colossal traffic for a definite protocol, unusually high level of loading of the net.

Different methods to track the packages are used to identify the origin of the attack when the address of the source has been substituted. It is possible to put these methods into practice when keeping the intermediate knots (in the route tables of routers) of data for passing packages.

To counteract the attack, new rules of filtration are applied, obtained at the stage of identifying of the source and attempts are made to track out and make harmless the attacking. The most advanced mechanisms for protection provide realization of all these stages

Different classification schemes of the protecting mechanisms have been developed to provide security against DDoS attacks. They structure the objective field of DDoS and make easy the search of ways of protection.

CERT organization (Computer Emergency Response Team, [2]) has developed a number of recommendations and requirements towards Internet users and providers in order to avoid DDoS attacks and to minimize their consequences.

2. MODELING OF SECURITY AGAINST ATTACKS OF THE TYPE DISTRIBUTED DENIAL OF SERVICE TYPE

The effective analysis of the attacks on the computer system and the reactions in answer require a multi-aspect modeling which analyses the structure of the computer network configuration, defining of the weak points, the possibilities for antagonistic attacks.

The multi-aspect model can be used as a base for attack in real time. The modeling of the answer can be used as an instrument, estimating analytically the security of networks in the real world. The model presents the components of the network, type of servers, automated places of work, routers and protecting walls, protocols and services.

Fig. 2 shows a version of a model of an attack in which the topology of the computer network and the configuration are modeled in separate tables, corresponding to a "Defining of Computer Network Type" component.

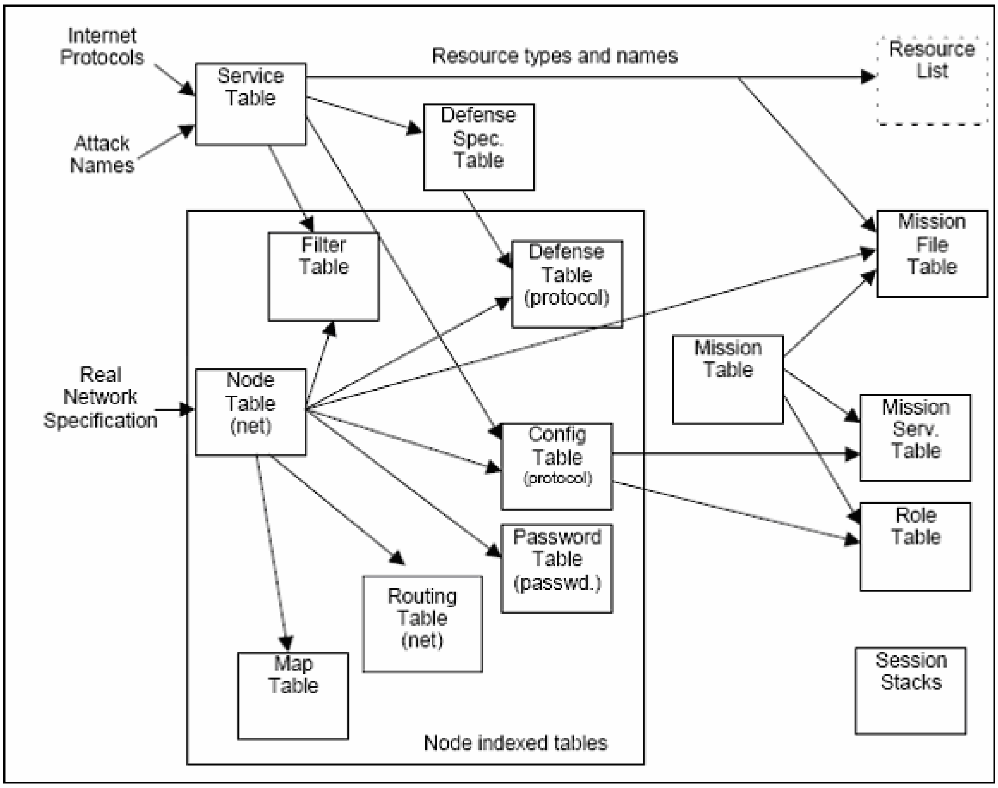

Fig. 2 Model of attack

The main elements are the following tables:

- Node Table includes determination of the specific character and peculiarities of the knots. A computer system is usually considered to be a knot but it might also be a legal entity in the present model or it might be a telephone system, an office, etc.

- Routing Table is generally used to define the route to a definite customer or to a specified address of the current knot.

- Configuration Table defines the configuration of the computer network for each protocol and each service which can be handled (software brand).

- Filter Table shows the rules, attributed to the routers in order to filter traffic.

- Defense and Defense Specification Tables store the conditions of defense, which are accessible for each knot.

- Password Table is used to present accounts and passwords.

- Map Table contains information about positions and dimensions of the elements in the computer network.

A team of scientists in SPIIRAS [10] are solving the problem of using multi-agent systems for modeling of complex antagonistic processes in order to protect information in computer networks. The originality of the obtained results is confirmed by the fact that, up to the present moment, the task, based on agents' technologies for modeling of computer counteraction against evil-minded people (hackers) and components for protection of information, has not been solved.

A set of different models (analytical, hybrid, imitative on the level of web packets, factual and measured) are used for research modeling.

A research prototype for modeling of distributed attacks Denial of Service type and mechanisms for defense against them, based on imitation of the level of web packets has been built. During the process of designing and realization of the agents, the following elements of the abstract architecture FIPA (Foundation for Intelligent Physical Agents) have been used: transport and network layer, communication language and agents' catalogue.

The basic result of the project is a creation of an integrated approach to build information protective systems, operating in aggressive, antagonistic surrounding. On the basis of ontology, a distributed base of knowledge about modeling of processes of computer counteracting have been developed, which is shown in Fig. 3. This ontology structures the main familiar types of attacks and their relationship. It includes a macro level at which it describes the structural relations of the multitude of attacks and a micro level, at which it describes the realization of the particular attacks in the form of sequences of attacking actions.

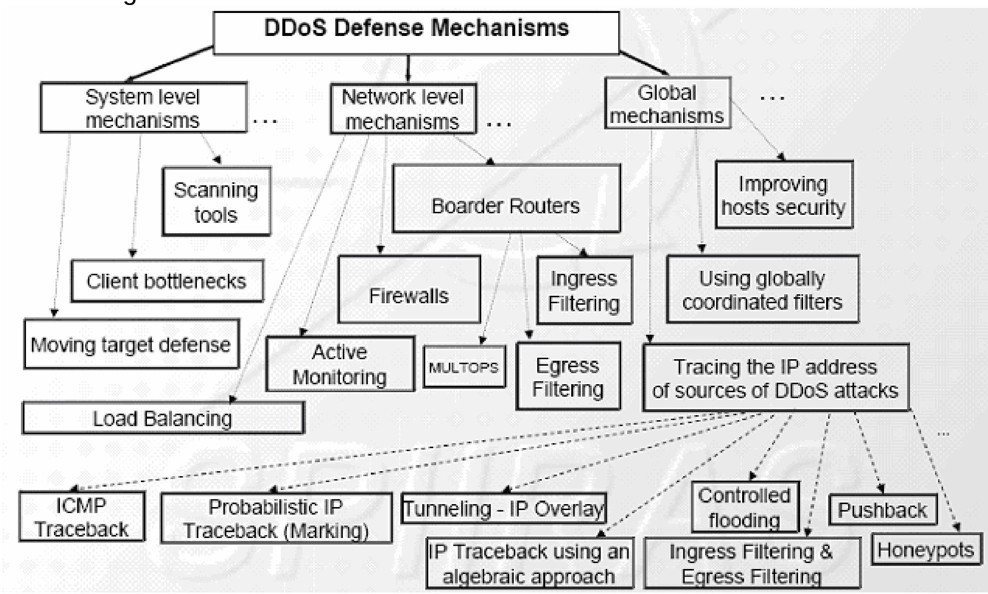

Fig 3. Ontology of the protective mechanism

There have been identified two main types of components of the attacking system: Daemon – an agent, directly stacking DoS and Master, - an agent, coordinating the other components of the system. The Daemons are able to attack at different modes of operation. The latter influences the ability of the defending team to protect, identify and block the attack and also to track and eliminate the agents of the attack. The Daemons can send attacking packets with varied intensity, to replace the address of the sender and to do all this at varied frequency.

3. PROGRAM AND APPARATUS MEANS OF PROTECTION AGAINST DISTRIBUTED DENIAL OF SERVICE TYPES OF ATTACKS

Of all the 4 elements, related to protection, the first one is the most important: providing of maximum good protection – warning of the attack. There have been put into practice a lot of attempts for automation of this process. A specialized solution is FloodGuard [12]. FloodGuard is a programming-apparatus complex. The functional chart of the product is shown in Fig. 4. This system has detectors on its protected walls, communicators and routers which constantly track the traffic and create its profile according to such characteristics as volume of packets of data, type, source, aim, etc.

When anomalies appear, the detector instantly raises an alarm and activates the actuators, sending them information about these anomalies, the source of attack, the size of the parasite traffic and the type of the sent packages. The actuators are placed in different segments of the net, on the routers, so that they can constantly follow the traffic and receive data about parasite packets. When parasite packets have been detected, the actuator immediately sends an alarm signal to the preceding module, standing on the way of the traffic before the actuator, the recommendations being activating of filters of the corresponding routers. In this way, the barrier for the avalanche of attacking data is raised and the harmful traffic is blocked temporally.

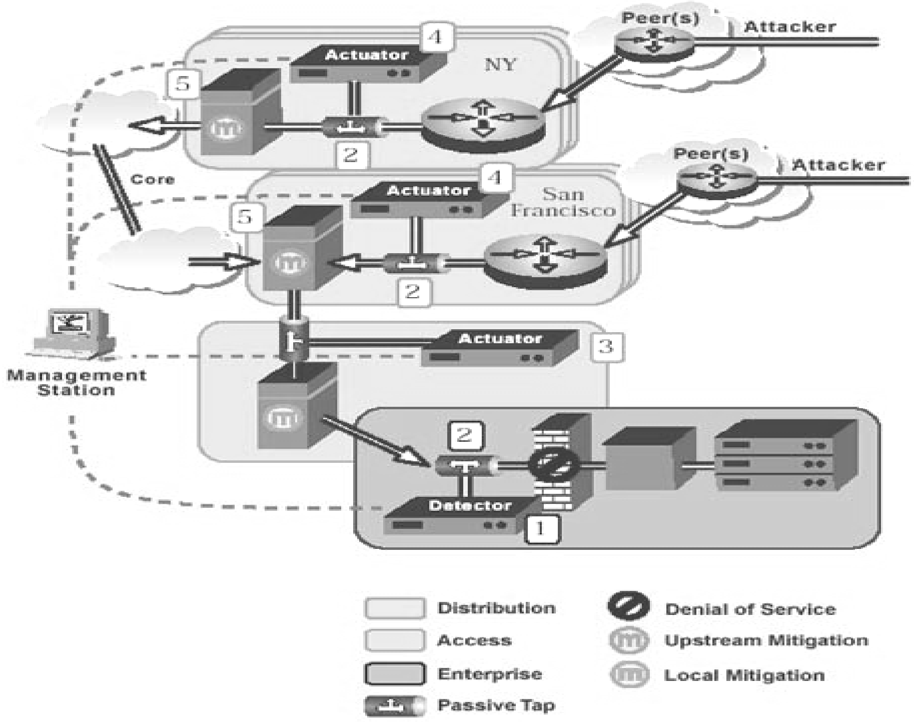

Fig. 4 Chart of operation of FloodGuard system

Intel offers a system, patented by the engineers David Putzolu and Todd Anderson, which modifies the routers in such a way that they automatically react to the alarm signal of the attacked computer [13]. It is supposed that the alarm signal contains a copy of the harmful packet. The routers immediately create a new profile (mask) and cut out all similar messages. If it is detected that the harmful message passes through the raised barrier then the alarm signal changes and the barrier is set in such a way that it totally blocks the parasite traffic. Intel's suggestion is technologically similar to the one, suggested by Reactive Networks two years ago. Intel offers a production of routers of a new generation with an Intel security module, which can identify and block the harmful traffic by themselves, while at the same time Reactive Networks sells its complex successfully.

4. CONCLUSION

The following conclusions can be drawn as a result of carried out studies and analysis:

Each server (web, ftp, dns) in practice is vulnerable to DDoS attacks. Efficient counteraction to attacks of such kind has not been discovered so far and specialists only give recommendations.

There is a tendency in the last years that people need less and less knowledge to realize attacks of such kind and consequently the system administrator should have more knowledge and skills to protect their net.

It is not easy to avoid DDoS attacks because of difficulties connected with their detection. Determining of the beginning of the attacks of such a kind, compared to the normal operation of the server, is defined by a great number of factors, which are unlikely to be identified by a single algorithm for protection.

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A FOREST-FIRE MODEL USING SPEED AND DIRECTION OF THE WIND

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***Abstract**: In this paper a forest-fire model using cellular automaton is described. The weak aspects of it are described. A new model with a set of input parameters based on cellular automaton is suggested. For this model six rules, which can be executed simultaneously are defined. A scheme of relational database storing the necessary information used by the model is presented. An example session of work with elaborated information system using the suggested model is presented. The future trends of research are described.*

***Key words:** cellular automata, forest-fire model.*

1. INTRODUCTION

This information system for growth prognosis of forest fire requires practical skills in the development of software. Also, skills to use the exist-ing new models or working out new ones are required for this aim. Using databases to store detailed information about forest areas and possibilities for operations with these data are a good premise to work out information systems with such functions. They must show the prognosticated growth of forest fire, in a researched area. The necessary information for this aim in database in suitable kind is collected and stored. In this paper the cellular automaton based model to prognosticate growth of forest fire is proposed. It is in conformity with the parameters – direction and speed of the wind. The values of the parameters are defined by the expert using the system. The developed information system, described in this paper, is only one of the possible examples to apply of proposed model. The three components: data, model to growth of forest fire and information system to calculated and show the results of the model work together. Together they can give an answer to the question: "Which area will the forest fire take place in any fu-ture moment?"

2. A FOREST-FIRE MODEL DESCRIBED WITH CELLULAR AUTOMATON

In 1989 Henley defined the model of forest fire as a cellular automaton on a grid with Ld cells. L is the side length of the grid and d is its dimension [3, 4]. A cell can be empty, occupied by a tree, or burning. In 1992 Drossel and Schwabl defined four rules which are executed simultaneously [1]:

1. A burning cell turns into an empty cell.
2. A tree will burn if at least one neighbor is burning.
3. A tree ignites with probability f even if no neighbor is burning.
4. All empty spaces are filled with trees with probability p.

The controlling parameter of the model is p/f which gives the average number of trees planted between two destructions by fire [2]. For the distri-bution of cells in clusters the next condition is necessary:

$f \ll p \ll Tsmax$,

where Tsmax is the time of burning the largest cluster [2]. A cluster is defined as a coherent set of cells, all of which have the same state. The cells are coherent if they can reach each other via the nearest neighbors. In most cases the four neighbor cells in horizontal or vertical are considered.

The first condition f << p, allows large structures to develop, while the second condition p << Tsmax, keeps trees from popping up alongside a cluster while burning.

The application of this model for the examination of growth of forest fire would not be successful. In [5] a classification depending on which part of trees the fire take places is made. It can be underground, on-ground and high. Depending on the speed of diffusion and height of flame, the forest fire is divided into: weak, middle and strong (see Tab. 1).

Tab. 1: Classification of forest fire according to [5].

Character of fire	**Speed of diffusion, [m/min]**		**Depth of burn, [m]**
	Over-ground	**High**	**Underground**
Weak	< 1	< 3	< 0,25
Middle	1 – 3	3 – 100	0,25 – 0,5
Strong	> 3	> 100	> 0,5

This classification does not take into consideration that model. Also the influence of other parameters such as direction and speed of wind are not taken into consideration. The different kind of inflammable materials and their location on the area is not mentioned. In this paper a model that requires the set of input parameters is proposed. The prognostication is made only on flat areas since the distribution of inflammable materials is well known. Only the application of the model on underground fires will be successful.

3. A FOREST-FIRE MODEL USED SPEED AND DIRECTION OF THE WIND

For p of numbers of inflammable materials, the linear speeds of diffusion of burning for a unit of time – v_p [m/min] are known. Also the time of full burning from a unit of area – t_p [min/m^2] is known. We examine a two-dimensional grid with cells – $c[m \times n]$. The cell can be empty, occupied or burning. Here under "occupied" we will understand filled with *p-th* inflammable material. The distribution of the fire materials in cells is known. We define six rules which can be executed simultaneously.

1. The user inputs or changes the values of the parameters used by the model – direction and speed of the wind.

2. With speed of the wind – $v_w = 0$ [m/sec] a cell will ignite, only if it is filled with an inflammable material, and if at least one neighbour is burning.

3. With speed of the wind – $v_w > 0$ the cell will ignite only if it is filled with an inflammable material and if at least one neighbour is burning form the direction of wind – d_w [0° – 360° degree].

4. With speed of the wind – $v_w > 0$ the values of the parameters v_p and t_p will be changed by εv_p and εt_p, depending on the quotients kv_p and kt_p.

5. From a burning cell fire will move into any neighbour (depending on rules 2 and 3), after long enough time – Δt_p, so that the fire cover the way s_c [m] equal to the size of the cell. I.e. (1) should be performed:

(1) $$\Delta t_p \geq (t - t_c)$$

where, t is the present moment, and t_c is the moment when the cell c_{ij} started burning; Δt_p is the necessary time for the fire to cover the way with speed v_p through a given cell filled with inflammable material p (2):

$$\Delta t_p = \frac{s_c}{v_p} \quad (2)$$

6. A burning cell filled with inflammable material p turns into a burned (empty) after time t_p.

The proposed model is characterize by the following:

- The values of v_p and t_p are given as "perfect" conditions. For v_p – without any wind. For t_p – with definite temperature and dampness both for inflammable material and environment.
- In the proposed model, the temperature and the dampness of the inflammable material, as well as environment are not considered. They will be researched in the next modification of the model with input of quotients for correction of the respective values.
- In the present model, the direction of wind is one of eight possible directions all of them at are distance from one another with an angle of 45°. d_w can accept one of the following values: 0° - east wind [E]; 45° - north-eastern wind [NE]; 90° - northern wind [N]; 135° - north-western wind [NW]; 180° - western wind [W]; 225° - south-western wind [SW]; 270° - southern wind [S]; 315° - south-eastern wind [SE].
- If the values of the parameters d_w – direction of the wind and v_w – speed of the wind for the given future moment are known, this can be changed even during the process.
- The quotients kv_p and kt_p are input for all inflammable materials. They show how much the values of v_p and t_p according to the speed of wind be corrected.

We

4. THE INFORMATION SYSTEM

This information system demonstrates the use of the proposed model. In a relational database information about the inflammable materials, the known areas and distribution of the inflammable materials in cells is stored (Fig. 1).

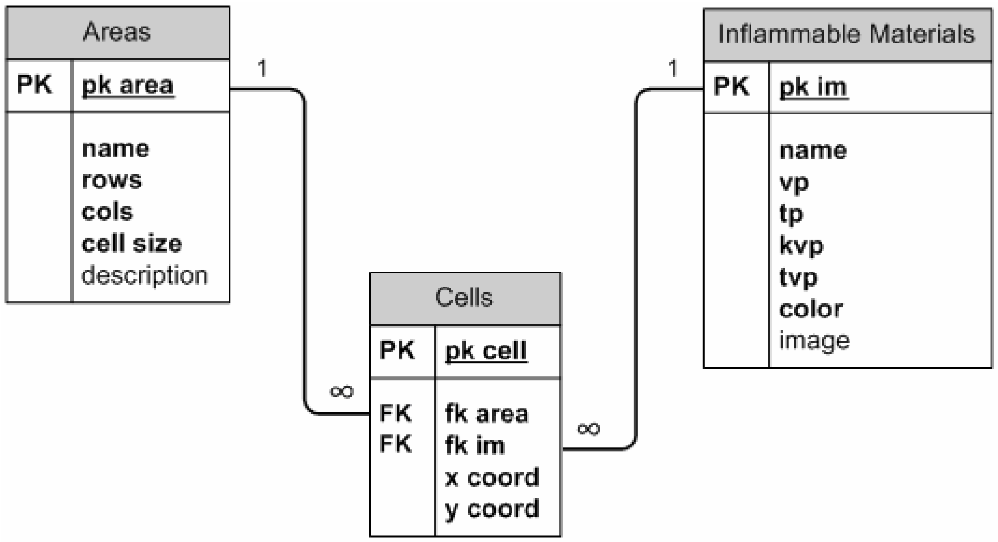

Fig. 1: Scheme of the relational database.

Information about the inflammable materials is stored in table "InflammableMaterials" (Tab. 2). Each inflammable material is described with a unique number, name, color, image and values of the parameters – v_p, t_p, kv_p и kt_p.

Tab. 2: Structure of table "Inflammable Materials".

No	Attribute	Domain	In PK[1]	In FK[2]	Null	Description
1	pk_fm	Identity	Yes	No	No	Primary key
2	name	Text(30)	No	No	No	Name
3	vp	Float	No	No	No	Value of v_p
4	tp	Float	No	No	No	Value of t_p
5	kvp	Float	No	No	No	Value of kv_p
6	tvp	Float	No	No	No	Value of tv_p
7	color	Text(20)	No	No	No	Linked color
8	image	BLOB	No	No	Yes	Image

[1] PK – Primary Key, [2]FK – Foreign Key

Information about the areas is stored in table "Areas" (Tab. 3). Each area is described with a unique number, name, number of cells in horizontal and vertical axes, size of the cell in this area and additional description.

Tab. 3: Structure of table "Areas".

No	Attribute	Domain	In PK	In FK	Null	Description
1	pk_area	Identity	Yes	No	No	Primary key
2	name	Text(20)	No	No	No	Name
3	rows	Integer	No	No	No	Number of rows
4	cols	Integer	No	No	No	Number of cols
5	cell_size	Small Integer	No	No	No	Cell size
6	description	MEMO	No	No	Yes	Description

Information about the cells is stored in table "Cells" (Tab. 4). Each cell is described with a unique number, a foreign key of the area that belongs to the cell, a foreign key of the inflammable material which fills the cell, x coordinate and y coordinate.

Tab. 4: Structure of table "Cells".

No	Attribute	Domain	In PK	In FK	Null	Description
1	pk_cell	Identity	Yes	No	No	Primary key
2	fk_area	Long Integer	No	Yes	No	Foreign Key
3	fk_fm	Long Integer	No	Yes	No	Foreign Key
4	x_coord	Integer	No	No	No	X coordinate
5	cell_size	Small Integer	No	No	No	Y coordinate

The cells which are not described in the database but belong to a given area are filled with non-inflammable material. When the user chooses an area in a two-dimensional dynamic array information about the cells form the database is loaded. The input parameters – direction and speed of the wind are given. Initial cell(s) in which burning has started are chosen and button "Start" is pressed (Fig. 2). A prognosticate process of burning diffusion begins. The burning cells are colored in red. When the time of burning of this cell is expires it is colored in black. When there are no burning cells a message of the end of the process is shown.

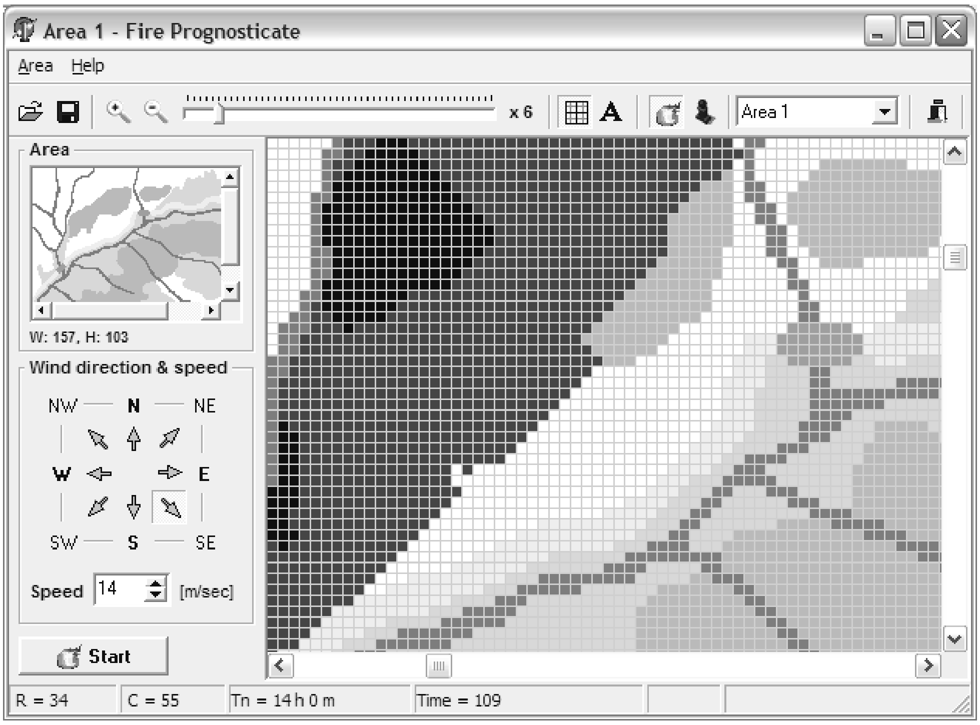

Fig. 2: Example session of work with the information system.

5. CONCLUSION

In this paper a model for fire growth prognosis, it is not compulsory a forest, but on an open-air area is described. For the model, definite input parameters are necessary. The parameters are inserted by an expert. The values of the input parameters can be changed even when the information system has started the process of prognostication. The advantages of the proposed model are the following: possibility to insert the type and properties of the inflammable materials; also the size of a cell from the grid. Disadvantages: input of the quotients for speed of burn correction and the time for the process of full burning of the inflammable materials depending on the speed of the wind. A model with all these aspects requires a serious future study.

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AUTENTIFICATION METHODS IN DISTRIBUTED SOFTWARE SYSTEM

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Abstract: *Software system for sensor simulations supports several authentication methods besides a simple password and allows some ways to keep safe access over networked connections. In this paper we discuss some distributed mechanisms to support authorisation and accounting. This paper presents the distributed model for authorisation and shows how the model can be used to support a wide range of authorisation and accounting mechanisms.*
Keyword: *Simulation and modelling, autentification, security*

INTRODUCTION

The Hall effect underlie in various semiconductor(particular Si and AIIIBV)magnetic sensors and Microsystems with great variety in their device constructions and omnidirective applicability[1]. This article proposes a new interpretation of the Hall effect, interpretation which eliminate one grave contradiction in existing model of this phenomenon[2]. The progress of the sensorics of magnetic field is illustrated also with newly invented parallel-field Hall effect[3].

Electronic authentication (e-authentication) is the process of establishing confidence in user identities electronically presented to an information system. E-authentication presents a technical challenge when this process involves the remote authentication of individual people over a network.

Protocol is a predetermined, ordered, unambiguous and complete series of steps, involving two or more participants, which is designed to achieve a particular goal. Security protocols, often called cryptographic protocols, are protocols which rely upon cryptography to provide security services across distributed systems. These protocols make use of cryptography in order to 'prevent or detect eavesdropping and cheating[4]

The paradigm of this document is that individuals are enrolled and undergo an identity proofing process in which their identity is bound to an authentication process. Thereafter, the modification of Kerberos framework are presented The authentication protocol allows an individual to demonstrate to a verifier that he has or knows the secret token, in a manner that protects the secret from compromise by different kinds of attacks. Higher authentication assurance levels require use of stronger tokens (harder to guess secrets) and better protection of the token from attacks.[6]

Authentication begins with registration. An applicant applies to a Registration Authority (RA) to become a subscriber of a Credential Service Provider (CSP) and, as a subscriber, is issued or registers a secret, called a token, and a credential that binds the token to a name and possibly other attributes that the RA has verified[7]. The token and credential may be used in subsequent authentication events. The subscriber's name may either be a verified name or a pseudonym. A verified name is associated with the identity of a real person and before an applicant can receive credentials or register a token associated with a verified name, he or she must demonstrate that the identity is a real identity, and that he or she is the person who is entitled to use that identity.[8]

DEFINITIONS

Principals: These are defined to be the legitimate entities involved in a particular protocol session. The principals should legitimately know or obtain the data transferred in the protocol. In some protocols the principals are assumed to be trustworthy and carry out their role in the protocol correctly, e.g. send messages correctly, not disclose shared keys to other principals. In other protocols, the principals do not trust each other to behave correctly. Principals may also be inter-changeably referred to as the participants or parties involved in the protocol. Examples of principals who may be involved in a protocol: vendors, customers, Trusted Third Parties and servers. In two way transaction principals are User *U* and Host *H*.[9]

Attacker: An attacker is defined to be a threat agent who tries to attack the protocol or use it in a maliciously manner. Where the protocol participants are assumed to be trustworthy, the attacker is not a legitimate party in the particular protocol run under consideration. They may, however, be a legitimate party in another run of the same or a different protocol on the same system. Where the protocol participants do not trust each other, the attacker may be a legitimate participant of the protocol.

Data: The data transferred in the protocol, as well as the data known by the participants at the start of the protocol. This includes components such as: orders, identities, credit card information, nonce's, timestamps and keys.

Checks: These are the actions, which may be carried out on message data received by a protocol participant. e.g. calculating and comparing hash values, checking the origin of a message etc.

AUTHENTICATION METHODS

Replay of old messages can be countered by using nonce's or timestamps. A nonce is information that is guaranteed fresh, that is, it has not appeared or been used before. Therefore, a reply that contains some function of a recently sent nonce should be believed timely because the reply could have been generated only after the nonce was sent. Perfect random numbers are good nonce candidates; however, their effectiveness is dependent upon the randomness that is practically achievable. Timestamps are values of a local clock. Their use requires at least some loose synchronisation of all local clocks, and hence their effectiveness is also somewhat restricted.

The address of initiator and responder and their nonce's or timestamps unique identifies authentication transaction. Furthermore the timestamps allow building explicit logs on every authentication transaction.

Class = <Uaddres, Haddres, TU, TH>

Hence if message includes this quadruple, it would be unique and cannot be exploited by attacker in another transaction.

A simple timestamp or nonce control is enabled.

The different classes of principals have strong restrictions on interclass transactions. Moreover some system attributes such as network address give us an opportunity to reject connection even before decryption stages.

We limit our discussion to the 4S authentication protocols and omit various administrative issues. 4S is a further extension of a Kerberos's design which is based on the use of a symmetric cryptosystem together with trusted third-party authentication servers. It is a refinement of ideas presented in [10]. Kerberos uses two main protocols.

The credential initialisation protocol authenticates user logging and installs initial tickets at the login host. A client uses the client-server authentication protocol to request services from a server.

(1)	*U* $\rightarrow$ *H*	: *U, A, Class, T1*
(2)	*H*	: check *Class;* if decryption fails, abort login
		: Retrieve A, U, Class, T1
(3)	UIS	: retrieve k{U,A,Class}_ _ and k4S__from database
		: generate new session key *k*
		: create ticket-granting ticket *ticket*
(4)	UIS	: Stamp{*A, T2*}if decryption fails, abort login
(5) UIS		:*kU*{ *U,A,Class, k, T1,T2, L* }
(6)	*U* $\rightarrow$ *H*	*: "Password?"*
(7)	*H* $\rightarrow$ *U*	*: passwd*
(8)	*H*	*:compute p = f(passwd)*
		:recover *k,ticket* by decrypting *kU*{ *U,A,Class, k, T1,T2, L,* } with *p*
		: if decryption fails, abort login; otherwise retain *k,tickTGS* and _
		: erase *passwd* from memory

Fig. 1: Credential Initialisation

The credential initialisation protocol uses UIS servers. Let *U* be a user who attempts to log into a host *H* . The protocol is specified in Figure 1.In step (1), user initiates login by entering his/her user name. In step (2), the login host *H* forwards the login request to a Kerberos server. In steps (3) and (4), the UIS server retrieves the user record of and returns a ticket-granting { *U,A,Class, k, T1,T2, L,*} to *H* , where *T1* is a timestamp and *L* is the ticket's lifetime. In steps (5) and (6), enters his/her password in response to _ 's prompt. In step (7), If *passwd* is not the valid password of , *ticket* would not be identical to *tickUIS* , and decryption in the last step would fail.5 Upon successful authentication, the host obtains a new session key *k* and a copy of *kU*{ *U,A,Class, k, T1,T2, L,*. The ticket-granting *ticket* is used to request server tickets from a UIS. Note that *ticket* is encrypted with tickUIS , the shared key of UIS.

(1)	*C* $\rightarrow$ UIS	:	*S,ticket,k{C,T1}*
(2)	UIS	:	recover *k* from *ticket* by decrypting with *kUIS*
		:	recover *T1* from {*C,T1*} by decrypting with *k*
		:	check *T1,* if decryption fails, abort login
		:	check *Class;* if decryption fails, abort login
		:	generate new session key *k"*
		:	create server ticket *tickS* = {*C,S,k',T1,L'*}*ks*
(3)	UIS $\rightarrow$ C		{*C,S,k',ticsS*}*k*
(4)	C		: recover *k'*, *trickS* by decrypting with *kS*
(5) *C* $\rightarrow$ S		:	*tickS, {C,T2}*
(6)	*S*	*:*	recover *kS* from *tickS* by decrypting with *k'*
		:	recover *T2* from {*C,T1*} by decrypting with *k*
		:	check *T2,* if decryption fails, abort login
(7)	*S* $\rightarrow$ *C*	*:*	*{T2 +1}k*

Fig. 2: Client-Server Authentication

Because a ticket is susceptible to interception or copying, it does not by itself constitute sufficient proof . Therefore, a principal presenting a ticket must also demonstrate knowledge of the session key named in the ticket. An authenticator (to be described) provides the demonstration. Figure 2 shows the protocol for a client _ to request network service from a server S.

T1 and *T2* are timestamps. In step (1), client *C* presents its ticket-granting ticket *ticket* to UIS to request a ticket for server. Knowledge of *k* is demonstrated using the authenticator {*C,T1*} In step (2), UIS decrypts *ticket*, recovers *k* , and uses it to verify the authenticator. If both step (2) decryption's are successful and *T1*is timely, UIS creates a ticket for server *S* and returns it to *C*. Holding *T1* and *C* repeats the authentication sequence with *k*

Thus, in step (5), tickS presents with *{C,T2}* and a new authenticator. In step (6), *S* performs verifications similar to those performed by UIS in step (2). Finally, step (7) assures *C* of the server's identity. This protocol requires synchronised local clocks for the verification of timestamps.

CONCLUSION

In literature, is given the description of the conjunction of authenticators usually named a combination of methods or principles, which is applied to the elimination of the threat of false acceptance. In practice we use the disjunction of authenticators, too.

A defence against false acceptance can weaken a defence against false rejection and vice versa.

A protocol development model can be used to avoid some of the pitfalls in the present development process. This structured model applies suitably adapted software and system safety engineering techniques in the phases of requirements, design, verification and validation, implementation, testing and maintenance. In particular, the use of requirement techniques in the development process helps to avoid the elementary but frequently occurring flaws and attacks on security protocols, as well as management and implementation errors. Heuristic modifications of kerberos framework improve reliability and performance of distributed systems. In 4S these improvements are especially effective due to advantages of Classed structure of replicants, owing to this fact connection requests could be rejected in early stages if authentication process.

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NEURAL NETS BASED MODELS FOR FORECASTING

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Abstract: *In most ecological systems forecasts of external requirement or prognosis of the future systems state are necessary to reach better management and control. In this article, we present two examples of forecasting applications based on artificial neural network. Forecasting tasks can be formulated as different classes of problems such as function approximation or interpolation state determination and classification in which neural network techniques can be applied. Trough critical analysis of this applications, our aim is to show the applicability of neural network techniques to forecasting problems, and trough analysis of these applications we put their constrains and limits and also their advantages and drawbacks under consideration.*
Keywords: *Artificial neural networks, forecasting, machine learning, statistical modelling*

INTRODUCTION

In many ecological systems, it is often necessary to have reliable forecasts of external demand for optimal management and control. Generally, the end-user's needs be classified in four categories depending on the relationship between the time constant of the underlying system and the forecasting time scale considered. For instant, in ecological context, the following types of forecast should be considered:

Long-term forecasts, corresponding to yearly evolution in ecological economic applications, are needed to design infrastructure and to plan investment. This type of forecast generally has an important ecological and economic impact because it provides information for high level decisions.

Middle-term forecasts are typically issued several months in advance in order to assist planning of production resources.

Short-term forecasts are often associated with a lead-time varying from one day to several hours. Such forecasts are an essential element for optimal control, decision aid in critical situations or detection of abnormal situations.

For monitoring and predicting of air pollution on short-range often is used the model of multilevel perception with right – and back- propagation, the methods of the analysis of temporary rows (ARMAX, ARX) and hybrid systems too [5,6].

Independent of that dispersion models are applied at the moment of pollution air the searching of the new ways of investigation progress because they don’t widespread in the case of changes on atmosphere’s dynamic conditions.

Real-time forecasting is concerned with sampling periods not exceeding a few minutes. In this case, the forecasting system performs automatically and safety becomes a critical issue.

Solving the following forecasting problems can provide answers to these specifications:

Trend detection – the question is to determine whether there exists a trend or not.

Regularity analysis – this occasion required the objective is to model the underlying mechanism producing.

Irregularity detection – exceptional events often correspond to crisis situations. The understanding of their underlying causes is often a difficult problem.

If this classification remains valid for a wide range of applications, the time intervals may depend on the application. Techniques used to tackle these different forecasting problems are generally based on the following approaches:

Deterministic modelling, which attempts to describe the physical laws involved by equations. High level prior knowledge and a detailed description of the system structure are needed.

Conceptual modelling provides a global representation of the system at a macroscopic level. The parameters do not necessarily have a physical interpretation.

Expert systems are based on expert knowledge about the considered phenomena; this knowledge is often encoded in the form of rules.

Statistical modelling: in this case, the lack of prior knowledge is compensated by the search for dependencies in empirical data.

Neural network models belong to this last category. They are powerful machines learning techniques, which are able to extract the most relevant features from large data sets. This is useful in real-world applications involving complex systems that must be analysed on the basis of very week prior knowledge.

It is also important to adopt a well-defined methodology in the development of applications using neural network techniques.

In this article, we shall present two applications in various domains such as forecasting of water demand and air pollution These applications have posed several kinds of forecasting problems and have required suitable neural networks based solutions. Through critical analysis of these applications, our aim is to illustrate the general methodology for applying neural networks in forecasting, to highlight their constrains and limits, and to show the advantages and drawbacks of this approach as compared to other approaches.

A NEURAL MODELS FOR FORECASTING

Problem of the water demand is forecasts of daily water consumption are needed to optimise resources for water supply. We make use of real data coming from water distribution networks of different kinds because the causality factors are different for different kinds of water consumption: urban, industrial, and agricultural.

The best results were obtained with four-layer perceptrons. Much four-layer architecture gives almost the same results [1,2]. So, the smallest one is generally used, in order to reduce the learning phase. The architectures used are 20-6-2-1 and 20-6-6-1. A comparison with an ARIMA model showed the superiority of the neural network-based approach.

The problem of Middle-term forecasts of daily traffic flow is used several months in advance, in order to assist the planning of road works and toll-gate management. Short-term forecasts of traffic flow with a time of one to three hours are essential in traffic control, the objective of which is to delay the formation of traffic jams by regulating the access to the network. The experiments were performed the traffic data observed at ST-Arnoult toll-gate, near Paris [3]. For middle term forecasting, learning was performed with ten years old data and the model was tested on the year 1992. Two networks have been trained, one for each direction (from and to Paris). The obtained results have shown that the neural network gives globally better forecasts than those made by an expert.

For short-term forecasting, learning was performed using the data of 1990 [4]. Predictions with a lead-time of one to tree hours were made for the year 1991. The obtained prediction results are generally satisfactory, although important errors can still occur in same situations. However, in this case, the uncertainly measure is generally higher. Additional data with more examples on the typical traffic situations will help to reduce this kind of prediction errors.

Neural networks base models have been proposed to solve two different traffic-forecasting problems. In the case of middle-term forecasting, it has been that a multi-leer back-propagation network can approximate correctly the complex relationships between daily traffic on the one hand, and calendar and school holiday information on the other hand

In the case of short-term forecasting, the role of the neural network was not to model an input-output relationship, but to create categories among the input data, and to generate prototypes representing each of these categories. The advantage such an approach is that adaptation to other locations of the highway network is very easy, because the model has only one control parameter. Unlike general-purpose neural-networks that perform as black boxes, interpretation or labelling of prototypes is possible. The estimation of uncertainly associated to each prediction is also as an interesting functionality by the end-users.

TOWARDS A METHODOLOGY FOR DEVELOPMENT OF PROJECT OF PREDICTING AIR POLLUTION USES CONNECTIONIST MODELS.

After the experiments performed, it appears clearly that the solution to a problem and the method to find it depends on the data. However, based on the accumulated experience, the following recommendations may be given as guidelines for future developments. Our approach to Neural Network based application is based on four phases: development of references methods, parameter estimation, performance evaluation, and implementation. Even if the learning phase is often regarded as the most important, the other phases also have some specificity due to the use of neural networks. Even if the first approach recommended by [7,8]. Given in most cases the best results, in some other applications the third one has proved more efficient. The determination of the inner architecture of the neural networks has been one of the most fruitful research areas of predicting air pollution uses connectionist models.

This model forms the formal results as a concentration of the polutions of medium or sort time prediction. The procedures are used many variation of the main equation.

The emphasis of this article is to find the relation between time for training ANN to achieve desired network and the topology of chosen ANN. This is discussed in context of data of air pollution agents and abiotic factors for town of Darmstad measured 2003 y.

In resent years, the interest of pollution problem imposes incentives to create flexible multicriterial oriented models and knowledge.

As an example of such a constructive method developed within this project, The new algorithm has been proposed to build a neural network by optimising the internal representation of the data. Once the architecture has been determined, some tuning of the parameters has done.

The architecture of the neural network is determined according to two different goals. The objective of the forecasting algorithms have to taken into account and a control of the complexity of the solutions have to be performed since a “non-parametric ” model is proposed. The first constraint determines mainly the number of inputs and outputs, while the second one deals with the number and connectivity of hidden cells. When the objective at time t is not only to forecast some quantity x(t+1) but also x(t+2), x(t+3), x(t+w) for some integer w>1, three different architectures can be used:

- a single multi-leered perseptron is used to forecast x(t+1) and the forecast value is then used as a new input to forecast x(t+2) , and etc.;
- a single multi-leered perseptron is used with as many outputs as values to forecast;

  the W multi-leered perseptron are trained one for each output.

Even if in some techniques such as this algorithm the two phases are performed simultaneously, some practical tricks are typically useful in parameter estimation. Among them, one can emphases the following ones:

- initialisation to improve and accelerate this optimisation phase; this can be using decision tree, functional expansion or prototypes.
- back propagation acceleration algorithms
- pruning methods (model selection) using, e.g. optimal brain damage or regularisation approaches such as proposed in [9,10]
- selecting the cost function to be minimised
- stopping criteria
- multiple trials. Choose the best one or use "committee based" decision methods such as the boosting algorithm.

This is only a first guideline of an expected methodology for the use networks but in order to be able to design it completely more theoretical work for a deeper understanding of neural networks is needed.

CONCLUSION

Neural network models allow building efficient forecasting applications. But the development method has to rely on sound statistical principles. A simple learning phase can take quite a long- time, and the determination of a relevant set of parameters may require many trails. Therefore, it is nor sure that a neural network application can be developed in short time than useful method, as it is currently claimed.

We pointed out that an independent cross-validation set had to be used to have a good stopping criterion. About the adaptive phase, we notice that there may exist an optimal adaptive step, or at least an “optimal step size region”. The consequence of such a result is that it is no longer interesting to test all the algorithms, but the relevant questions become when to use such algorithm? In other words, the important point is to determinate on what kind of problem such or such algorithm is expected to yield good result. It is the reason why it is important to be able to classify the different kinds of problems according to some criterion to be determined.

The examples presented in this paper have demonstrated that these are no neural network paradigm suitable for all kinds of forecasting problems. For each problem, detailed analysis of domain data and the acquisition of prior knowledge are necessary to find a suitable connectionist model. Although multi-layer back-propagation networks are the most commonly use, training from scratch with all possible inputs and all available raw data has often proved ineffective. According to our experience, the introduction of prior knowledge in input selection input encoding or architecture determination is often very useful, especially when the available domain data is limited. Another important point is that, in practice, users often need some explanation or indication of the uncertainly associated to a prediction far making their decisions. From this point of view, general-purpose black-box network models are not as convincing as networks having interpretation possibility.

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SEPARABLE AND DOMINATING SETS OF VARIABLES FOR THE FUNCTIONS

K. Chimev, Pl. Tchimev

(Plenary report)

Abstract: *We discuss some structural properties of the functions, with respect to their separable and dominating sets of variables.*

Theorem 1 If the set D is dominating of the set M, $M \cap D = \emptyset$, for the function $f(x_1, \dots, x_n)$, $n \geq 3$, which depends essentially on n variables, then for every $m \in \{2, \dots, n\}$ and for every variable $x_i \in M$, there exists the set S, $|S| = m$, such that S is separable for f, $x_i \in S$ and $S \cap D \neq \emptyset$.

Definition 1 We say that the function $f(x)$ depends essentially on x, if $f(x)$ takes at least two values.

We say that the function $f(x_1, \dots, x_n)$, $n \geq 2$, depends essentially on x_i, $1 \leq i \leq n$, if there exist such values of the rest of the variables, that after replacing them in f we obtain a function which depends essentially on x_i.

The variables on which the function f depends essentially will be called essential variables for f, [1 - 7, 27 - 31, 36 - 45].

By $Ess(f)$ we denote the set of all essential variables for the function f.

The set of all functions which depend essentially exactly on n variables will be denoted by $F(n)$.

Definition 2 If $f \in F(n), n \geq 1$, and $\emptyset \neq M \subseteq Ess(f)$, then the variable $x \in M$ will be called *strongly essential* for f, with respect to M, if there exists a value c for x, such that

$$M \setminus \{x\} \subseteq Ess(f(x = c)), \quad [3 - 7, 30 - 31, 36 - 41].$$

The set of all strongly essential variables for the function f, with respect to M, will be denoted by $Ess^*(f, M)$.

Definition 3 If $f \in F(n), n \geq 1$, then the variable x will be called strongly essential for f, if it is strongly essential for f, with respect to $Ess(f)$.

The set of all strongly essential variables for the function f will be denoted by $Ess^*(f)$.

Definition 4 If $f \in F(n), n \geq 2$, and $\emptyset \neq M_1 \subseteq Ess(f)$, $\emptyset \neq M_2 \subseteq Ess(f)$, $M_1 \cap M_2 = \emptyset$, then we say that M_1 is *separable* for

f, with respect to M_2, if for the variables from M_2 there exist values, such that after replacing variables with these value, a new function, which is obtained from f, depends essentially of all variables, which belong to M_1.

If $M_2 = \emptyset$, then we assume that each subset of $Ess(f)$ is separable for f, with respect to M_2.

If $f \in F(n), n \geq 1$, and $\emptyset \neq M \subseteq Ess(f)$, then by $Sep(f, M)$ will be denoted the set of all non-empty sets, which are separable for f, with respect to M.

Definition 5 If $f \in F(n), n \geq 1$, then we say that the set M, $M \subseteq Ess(f)$, is separable for f, if M is separable for f, with respect to $Ess(f) \setminus M$.

The set of all non-empty sets which are separable for the function f will be denoted by $Sep(f)$.

When we say that the m-tuple $(x_1, \dots, x_m)$, is separable for the function $f \in F(n)$, $n \geq m$, we will know that the set $\{x_1, \dots, x_m\}$ is separable for f.

Definition 6 Let $f \in F(n), n \geq 3$. The set $D = \{x_1, \dots, x_m\}$, $\emptyset \neq D \subset Ess(f)$, is called dominating of the set M, $\emptyset \neq M \subset Ess(f)$, for f, if there exist values $c_1, \dots, c_m$ for $x_1, \dots, x_m$, such that

$$Ess(f_1) \cap M = \emptyset, \quad f_1 = f(x_1 = c_1, \dots, x_m = c_m).$$

We will say that the variable x_i is dominating of the set M, for the function f, if the set $\{x_i\}$ is dominating of the set M for f.

Definition 7 Let $f \in F(n), n \geq 3$ and D is dominating of the set M for the f. The set D is called minimal dominating of the set M, for f, if for every set $D_1 = \{x_1, \dots, x_p\}$, $\emptyset \neq D_1 \subset D$, and for every values $c_1, \dots, c_p$ for $x_1, \dots, x_p$,

$$M \cap Ess(f(x_1 = c_1, \dots, x_p = c_p)) \neq \emptyset, \quad [42].$$

Definition 8 The sets D_1 and D_2 are mutually dominating for the function f, if D_1 is dominating of the D_2 for f, and D_2 is dominating of the D_1 for f, [42].

Theorem 2 If $f \in F(n), n \geq 2$ and $M_1 \in Sep(f, M_2)$, $\emptyset \neq M_2$, $M_2 \cap M_1 = \emptyset$, then at least one of the variables from M_2 is strongly essential for f, with respect to $M_1 \cup M_2$.

Theorem 3 If $f \in F(n), n \geq 2$ and $M \subseteq Ess(f)$, $|M| \geq 2$, then for every $x_i \in M$, there exists a variable $x_j \in M \setminus \{x_i\}$, which is strongly essential for f, with respect to M.

Theorem 4 If $f \in F(n), n \geq 2$ and $M \subseteq Ess(f)$, $|M| \geq 2$, then at least two variables from M are strongly essential for f, with respect to M.

Theorem 5 Each function, which depends essentially on at least two variables, has at least two strongly essential variables.

Theorem 6 If the set D is dominating of the set M, $M \cap D = \emptyset$, for $f(x_1, \ldots, x_n) \in F(n), n \geq 3$ then for every $x_i \in M$, there exists $x_j \in D$, such that $\{x_i, x_j\} \in Sep(f)$.

Theorem 7 If the set D is dominating of the set M, $M \cap D = \emptyset$, for $f(x_1, \ldots, x_n) \in F(n), n \geq 3$ then for every $m \in \{2, \ldots, n\}$ and for every variable $x_i \in M$, there exists the set S, $|S| = m$, such that $S \in Sep(f)$, $x_i \in S$ and $S \cap D \neq \emptyset$.

Corollary 1 If the sets D_1 and D_2, $D_1 \cap D_2 = \emptyset$, are mutually dominating for the function $f(x_1, \ldots, x_n) \in F(n), n \geq 3$ then for every $m \in \{2, \ldots, n\}$ and for every variable $x_i \in D_1(D_2)$, there exists the set S, $|S| = m$, such that $S \in Sep(f)$, $x_i \in S$ and $S \cap D_2 \neq \emptyset (S \cap D_1 \neq \emptyset)$.

Theorem 8 If the set D is minimal dominating of the set M, $M \cap D = \emptyset$, for the function $f(x_1, \ldots, x_n) \in F(n), n \geq 3$ then for every $x_i \in D$, there exists $x_j \in M$, such that $\{x_i, x_j\} \in Sep(f)$, and for every $x_l \in M$, there exists $x_p \in D$, such that $\{x_l, x_p\} \in Sep(f)$.

Theorem 9 If the set D is minimal dominating of the set M, $M \cap D = \emptyset$, for $f(x_1, \ldots, x_n) \in F(n), n \geq 3$ and $m \in \{2, \ldots, n\}$, then:

a) for every $x_i \in D$, there exists a set S, $|S| = m$, such that $S \in Sep(f)$, $x_i \in S$ and $S \cap M \neq \emptyset$;

b) for every $x_i \in M$, there exists a set S, $|S| = m$, such that $S \in Sep(f)$, $x_i \in S$ and $S \cap D \neq \emptyset$.

Is the following assertion true:

If the set D is minimal dominating of the set M, $|M| \geq 2$, $M \cap D = \emptyset$, for the function $f(x_1, \ldots, x_n) \in F(n), n \geq 3$ then there exists $x_i \in D$, such that for every $x_j \in M$, $\{x_i, x_j\} \in Sep(f)$.

The answer is negative. This may be observed by the following example.

Let

$$f = x_1 y_1 y_2 + \bar{x}_1 z_1 z_2 \pmod 2$$

and

$$D = \{y_1, z_1\}, \quad M = \{y_2, z_2\}.$$

The set D is minimal dominating of the set M, for the function f, but $\{y_1, z_2\} \notin Sep(f)$ and $\{z_1, y_2\} \notin Sep(f)$.

Definition 9 The unoriented graph with vertices the all essential variables for the function $f(x_1, \ldots, x_n) \in F(n)$, and with edges the set of the all separable pairs for $f(x_1, \ldots, x_n)$, will be called graph of the function $f(x_1, \ldots, x_n)$.

Let we denote by $F^*(n)$ the set of all functions $f(x_1, \ldots, x_n) \in F(n), n \geq 3$, for which

$$Ess(f) = \bigcup_1^p M_i, \quad p \geq 3, \quad M_i \neq \emptyset, \quad i, j \in \{1, \ldots, p\}, \quad i \neq j$$

and for every two variables $x_l \in M_i$ and $x_m \in M_j$, $i \neq j$, $i, j \in \{2, \ldots, p\}$ the set $\{x_l, x_m\} \notin Sep(f)$.

Theorem 10 If $f(x_1, \ldots, x_n) \in F^*(n), n \geq 3$, then:

a) for every $l \in \{2, \ldots, p\}$, the set M_1 is dominating for $\bigcup_i M_i$, $i \in \{2, \ldots, p\} \setminus \{l\}$;

b) for every $l \in \{2, \ldots, p\}$, the set $Ess(f) \setminus M_l$ is separable for f.

Theorem 11 If $f(x_1, \ldots, x_n) \in F^*(n)$, then for every $m \in \{2, \ldots, n\}$ and for every $x_i \in \bigcup_2^p M_i$ there exists a set S, such that

$$S \in Sep(f), \quad |S| = m, \quad x_i \in S, \quad \text{and} \quad S \cap M_1 \neq \emptyset.$$

Theorem 12 If $f(x_1, \ldots, x_n) \in F^*(n)$ and $p \geq 4$, then the subgraph of the function f, with vertices the variables of M_1 is connected.

Theorem 13 If $f(x_1, \ldots, x_n) \in F^*(n) = M_1 \cup M_2 \cup M_3$ and the subgraph of the function f, with vertices the variables of M_1 is not connected, then for every $i \in \{2, 3\}$ the subgraph of the function f, with vertices the variables of M_i is connected.

Definition 10 Let $f \in F(n), n \geq 3$, and D is dominating of the set M, $\emptyset \neq M \subset Ess(f)$, for f. The set D is called direct dominating of the set M, if for every set M_1, $M \subset M_1 \subset Ess(f)$, the set D is not dominating of the set M_1.

Theorem 14 If $f \in F(n), n \geq 3$, and D is direct dominating of the set M, $M \cap D = \emptyset$, for f, then at least one of the variables from $M \cup D$ is strongly essential for f.

Theorem 15 If $f \in F(n), n \geq 3$, and D, $D \in Sep(f)$ is direct dominating of the set M, $M \cap D = \emptyset$, for f, then at least one of the variables from M is strongly essential for f.

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CLASSES OF SUBSETS OF X^n

Aleksa Malcheski

Abstract. *The notion of n-norm is introduced in [2], as generalization of the notion of 2-norm, introduced in [3] by S.Gahler. In [1] is introduced the notion of n-semi norm as generalization of the notion of 2-seminorm. Equivalent definition of 2-norm is given in [4]. In this paper, using similar technique as in [4], we consider special classes of subsets of* X^n *, which are important for the characterization of any n-semi norm.*

1. INTRODUCTION

Let X be a vector space over a field $\mathbf{\Phi}$. We will generalize the notions of absorbing, balanced and convex set and introduce the notion of invariant subspace of X^n. For these sets, we will prove one property of n-semi norms.

We begin with the definition of n-semi norm, given in [1].

Definition 1. Let X be a vector space over the field $\mathbf{R}$ and let $p: X^n \to \mathbf{R}$ be a mapping which satisfies the following conditions:

$a)$ $p(x_1, x_2, ..., x_n) = 0$, for every linearly dependant set $\{x_1, ..., x_n\}$.

$b)$ $p(x_1, ..., x_n) = p(x_{\pi(1)}, ..., x_{\pi(n)})$, for every permutation $\pi : \{1, ..., n\} \to \{1, ..., n\}$.

$c)$ $p(\alpha x_1, x_2, ..., x_n) = |\alpha| p(x_1, x_2, ..., x_n)$, for any scalar α and for all $x_1, ..., x_n \in X$,

$d)$ $p(x_1 + x_1^{'}, x_2, ..., x_n) \le p(x_1, x_2, ..., x_n) + p(x_1^{'}, x_2, ..., x_n)$, for all $x_1, x_1^{'}, x_2 ..., x_n \in X$.

The function $p: X^n \to \mathbf{R}$ which satisfies the conditions $a) - d)$ is called n-semi norm, and the ordered pair (X, p) is called n-semi normed space.

The following definition of n-norm is introduced in [2].

Definition 2. Let X be a vector space over the field of real numbers with $\dim X \geq n$. The function $\|\cdot,\dots,\cdot\| : X^n \to \mathbf{R}$ which satisfies the conditions:

$(N1)$ $\|x_1,\dots,x_n\| \geq 0$ i $\|x_1,\dots,x_n\| = 0$ if and only if the set $\{x_1,\dots,x_n\}$ is linearly dependant,

$(N2)$ $\|x_1,\dots,x_n\| = \|x_{\pi(1)},\dots,x_{\pi(n)}\|$, for all $x_1,\dots,x_n \in X$ and for every permutation $\pi : \{1,\dots,n\} \to \{1,\dots,n\}$,

$(N3)$ $\|x_1,\dots,x_{i-1},\alpha x_i, x_{i+1},\dots,x_n\| = \alpha \|x_1,\dots,x_{i-1},x_i,x_{i+1},\dots,x_n\|$, for all $x_1,\dots,x_n \in X$ and for every scalar $\alpha \in \mathbf{R}$ and any $i = 1,2,\dots,n$,

$(N4)$ $\|x_1 + x_1', x_2,\dots,x_n\| \leq \|x_1,x_2,\dots,x_n\| + \|x_1',x_2,\dots,x_n\|$, for all $x_1, x_1', x_2,\dots,x_n \in X$,

is called n-norm of the vector space X, and the ordered pair $(X, \|\cdot,\dots,\cdot\|)$ is called n-normed real space.

In $[4]$ is given equivalent definition of 2-norm, where the axioms of 2-norm are replaced with the following axioms, equivalent with the once given in the definition 2, for n=2.

$(P1)$ If $\|x,y\| = 0$ then the set of vectors $\{x,y\}$ is linearly dependant.

$(P2)$ $\|A(x,y)^T\| = |\det A| \cdot \|x,y\|$, for all $x,y \in X$ and $A \in M_2(\mathbf{R})$.

$(P3)$ $\|x + x', y\| \leq \|x,y\| + \|x',y\|$, for all $x, x', y \in X$.

Using similar technique as in $[4]$, we get new classes of subsets, which will be considered in the subsequent part.

2. n-ballanced, n-absorbing and n-invariant subspaces of X^n

Definition 1. The set $V \subseteq X^n$ is n-balanced if for any matrix $A \in M_n(\mathbf{\Phi})$ with $|\det A| \leq 1$ holds $AV^T \subseteq V$, where

$$V^T = \{(x_1,\dots,x_n)^T \mid (x_1,\dots,x_n) \in V\}.$$

Lemma 1. The intersection of arbitrary family of n-ballanced sets is n-ballanced set.

Proof. Let $V_\gamma, \gamma \in \Gamma$ be a family of n-ballanced sets and $V = \bigcap_{\gamma\in\Gamma} V_\gamma$. For arbitrary $A \in M_n(\mathbf{\Phi})$ holds $AV_\gamma \subseteq V_\gamma^T$, $\gamma \in \Gamma$, since $V_\gamma, \gamma \in \Gamma$ are n-ballanced sets. Acording to this, we have
$AV^T = A(\bigcap_{\gamma\in\Gamma} V_\gamma)^T = \bigcap_{\gamma\in\Gamma} AV_\gamma^T \subseteq \bigcap_{\gamma\in\Gamma} V_\gamma$, which means that $V = \bigcap_{\gamma\in\Gamma} V_\gamma$ is n-ballanced set.

Definition 2. The subset $S \subseteq X^n$ is n-convex if for any $\left(x_1,\ldots,x_{i-1},x_i,\ldots,x_n\right)$, $(x_1,\ldots,x_{i-1},x_i^{'},\ldots,x_n) \in S$ and for each $t \in [0,1]$ holds $(x_1,\ldots,x_{i-1}, tx_i + (1-t)x_i^{'},\ldots,x_n) \in S$.

Lemma 2. The intersection of arbitrary family of n-convex sets is n-convex set.

Proof. Let $S_\gamma, \gamma \in \Gamma$ be a family of n-convex sets and $S = \bigcap_{\gamma\in\Gamma} S_\gamma$. For $\left(x_1,\ldots,x_{i-1},x_i,\ldots,x_n\right)$, $(x_1,\ldots,x_{i-1},x_i^{'},\ldots,x_n) \in S$ we get

$$\left(x_1,\ldots,x_{i-1},x_i,\ldots,x_n\right),\ (x_1,\ldots,x_{i-1},x_i^{'},\ldots,x_n) \in S_\gamma, \gamma \in \Gamma .$$

Since $S_\gamma, \gamma \in \Gamma$ are n-convex sets, for any $t \in [0\ ,1]$ we have that

$$(x_1,\ldots,x_{i-1},tx_i + \left(1-t\right)x_i^{'},\ldots,x_n) \in S_\gamma, \gamma \in \Gamma$$

i.e.

$$(x_1,\ldots,x_{i-1},tx_i + \left(1-t\right)x_i^{'},\ldots,x_n) \in \bigcap_{\gamma\in\Gamma} S_\gamma = S ,$$

which proves the n-convexity of S.

Definition 3. The subset $P \subseteq X^n$ is called *n*-invariant if for each matrix $A \in M_n(\mathbf{\Phi})$ with $\det A = 1$ holds $AP^T \subseteq P$, where

$$P^T = \{(x_1,\ldots,x_n)^T \mid (x_1,\ldots,x_n) \in P\}.$$

Lemma 3. The intersection of arbitrary family of n-invariant sets is n-invariant set.

Proof. Let $P_\gamma, \gamma \in \Gamma$ be arbitrary family of n-invariant sets. Then, for any matrix $A \in M_n(\mathbf{\Phi})$ with $\det A = 1$, we have that $AP_\gamma^T \subseteq P_\gamma$. From

$$AP^T = A(\bigcap_{\gamma\in\Gamma} P_\gamma)^T = A(\bigcap_{\gamma\in\Gamma} P_\gamma^T) = \bigcap_{\gamma\in\Gamma} AP_\gamma^T \subseteq \bigcap_{\gamma\in\Gamma} P_\gamma = P$$

we conclude that $P = \bigcap_{\gamma\in\Gamma} P_\gamma$ is n-invariant space.

Definition 4. The subset $W \subseteq X^n$ is called n-absorbing if for each $\left(x_1, \ldots,x_n \right) \in X^n$ it exists $A_{x_1,\ldots,x_n} \in M_n(\mathbf{\Phi})$ with $\det A_{x_1,\ldots,x_n} > 0$, so that

$$\left(x_1 ,\ldots, x_n\right) \in A_{x_1,\ldots,x_n} W^T \text{, i.e. } \left(A_{x_1,\ldots,x_n}\right)^{-1} \left(x_1 ,\ldots, x_n\right)^T \in W .$$

We will consider the set $\Delta_n \subseteq X^n$ that consists of all elements $\left(x_1, \ldots,x_n \right) \in X^n$ for which the set $\left\{ x_1, \ldots,x_n \right\}$ is linearly dependant.

Lemma 4. If $W \subseteq X^n$ is n-absorbing and n-invariant set then $\Delta_n \subseteq W$.

Proof. For arbitrary $\left(x_1,\ldots,x_n \right) \in \Delta_n$, the set $\left\{ x_1,\ldots,x_n \right\}$ is linearly dependant. This means that there exists $i \in \left\{1,2,\ldots,n\right\}$ and scalars $\alpha_1,\ldots,\alpha_{i-1},\alpha_{i+1},\ldots,\alpha_n$ so that $x_i = \sum_{j \neq i, j=1}^{n} \alpha_j x_j$. Since W is n-absorbing set, there is $B \in M_n(\Phi)$ with $\det B > 0$ such that

$$B^{-1}(x_1,...,x_{i-1},0,x_{i+1},...,x_n)^T \in W.$$

Let $\beta = \sqrt[n]{\det B}$. For the matrix

$$A = \begin{bmatrix} \beta & 0 & 0 & ... & 0 & 0 & 0 & ... & 0 \\ 0 & \beta & 0 & ... & 0 & 0 & 0 & ... & 0 \\ 0 & 0 & \beta & ... & 0 & 0 & 0 & ... & 0 \\ ... & ... & ... & ... & ... & ... & ... & ... & ... \\ 0 & 0 & 0 & ... & \beta & 0 & 0 & ... & 0 \\ \alpha_1\beta & \alpha_2\beta & \alpha_3\beta & ... & \alpha_{i-1}\beta & \frac{1}{\beta^{n-1}} & \alpha_{i+1}\beta & ... & \alpha_n\beta \\ 0 & 0 & 0 & ... & 0 & 0 & \beta & ... & 0 \\ ... & ... & ... & ... & ... & ... & ... & ... & ... \\ 0 & 0 & 0 & ... & 0 & 0 & 0 & ... & \beta \end{bmatrix},$$

$\det A = 1$. Since $\det B > 0$, we have $\det B^{-1} = \frac{1}{\det B} > 0$. So, for $C = B^{-1}$, holds $C(x_1,...,x_{i-1},0,x_{i+1},...,x_n)^T \in W$. For the matrix $\frac{1}{\beta}C^{-1}$ we have

$$\det\left(\frac{1}{\beta}C^{-1}\right) = \frac{1}{\det B}\det C^{-1} = \frac{1}{\det B}\det B = 1.$$

Using that W is n-invariant set, we get that

$$\left(\frac{1}{\beta}C^{-1}\right)\left(C(x_1,...,x_{i-1},0,x_{i+1},...,x_n)^T\right)^T \in W$$

which is equivalent to $\left(\frac{1}{\beta}C^{-1}C\right)(x_1,...,x_{i-1},0,x_{i+1},...,x_n)^T \in W$, i.e.

$$\left(\frac{1}{\beta}E\right)(x_1,...,x_{i-1},0,x_{i+1},...,x_n)^T \in W$$

i.e.

$$\left(\frac{1}{\beta}x_1,...,\frac{1}{\beta}x_{i-1},0,\frac{1}{\beta}x_{i+1},...,\frac{1}{\beta}x_n\right)^T \in W.$$

Since the matrix A has $\det A = 1$, and again, using that W is *n*-invariant, we get that:

$\left(x_1,...,x_{i-1},x_i,x_{i+1},...,x_n \right) = A\left(\frac{1}{\beta}x_1,...,\frac{1}{\beta}x_{i-1},0,\frac{1}{\beta}x_{i+1},...,\frac{1}{\beta}x_n \right)^T \in W$ Finally $\Delta_n \subset W$.

Lemma 5. Every *n*-balanced set is n-invariant set.

Proof. Let V is *n*-balanced set and $A \in M_n(\mathbf{R})$ with $\det A = 1$. Then $AV^T \subseteq V$. This means that V is n-invariant set.

3. n -semi norms and classes of subsets of X^n

Theorem 1. Let $p : X^n \to \mathbf{R}$ be an *n*-semi norm. Then the set $F = \left\{ \left(x_1,...,x_n \right) \mid \ p\left(x_1,...,x_n \right) < 1 \right\}$ is *n*-convex, *n*-absorbing and *n*-balanced.

Proof. Let $\left(x_1,...,x_n \right) \in F$ i.e. $p\left(x_1,...,x_n \right) < 1$ and $A \in M_n(\mathbf{R})$ with $\left| \det A \right| \le 1$. Using the definition of n-semi norm, we have that

$$p\left(A\left(x_1,...,x_n \right)^T \right) = \left| \det A \right| p\left(x_1,...,x_n \right) \le 1 \cdot p\left(x_1,...,x_n \right) < 1.$$

According to this, $A\left(x_1,...,x_n \right)^T \in F$. Since $\left(x_1,...,x_n \right) \in F$ is arbitrary, we have that $AF^T \subseteq F$. Thus the set F is n-balanced set.

Let $\left(x_1,...,x_n \right) \in X^n$ is arbitrary chosen element. Using the definition of n-semi norm, $p\left(x_1,...,x_n \right) \ge 0$. There exists a matrix $B \in M_n\left(\mathbf{\Phi} \right)$ such that $\det B > p(x_1,x_2,... ,x_n) \ge 0$. Then

$$p\left(B^{-1}\left(x_1,x_2,...,x_n \right)^T \right) = |\det B^{-1}|\, p\left(x_1,x_2,...,x_n \right) = \frac{1}{\det B} p(x_1,x_2,...,x_n) < 1$$

This means that $B^{-1}\left(x_1,x_2,...,x_n \right)^T \in F$, i.e. $\left(x_1,x_2,...,x_n \right) \in BF^T$. Since $\left(x_1,...,x_n \right) \in X^n$ is arbitrary, we get that F is absorbing set.

Let $\left(x_1,...,x_{i-1}, x_i ,x_{i+1},...,x_n \right), \left(x_1,...,x_{i-1}, x_i' ,x_{i+1},...,x_n \right) \in F$ and $t \in \left[0,1 \right]$ are arbitrary chosen elements. Using the condition $p(x_1,...,x_{i-1}, x_i ,x_{i+1},...,x_n), p(x_1,...,x_{i-1}, x_i' ,x_{i+1},...,x_n) < 1$ and the definition of n-semi norm, we get

$$p\left(x_1,...,x_{i-1}, tx_i + (1-t)x_i^{'}, x_{i+1},...,x_n \right) \le$$
$$\le p\left(x_1,...,x_{i-1}, tx_i, x_{i+1},...,x_n \right) + p\left(x_1,...,x_{i-1}, (1-t)x_i^{'}, x_{i+1},...,x_n \right)$$
$$= \left(\det A \right) p\left(x_1,...,x_{i-1}, x_i, x_{i+1},...,x_n \right) + \left(\det B \right) p\left(x_1,...,x_{i-1}, x_i^{'}, x_{i+1},...,x_n \right)$$ where
$$= tp\left(x_1,...,x_{i-1}, x_i, x_{i+1},...,x_n \right) + \left(1-t \right) p\left(x_1,...,x_{i-1}, x_i^{'}, x_{i+1},...,x_n \right) < 1$$

A is elementary matrix in whose diagonal on the *i*-th row is the number t , and B is as A with $1-t$ instead of t .

So, $\left(x_1,...,x_{i-1}, tx_i + (1-t)x_i^{'}, x_{i+1},...,x_n \right) \in F$, and using the arbitrariness of $\left(x_1,...,x_{i-1}, x_i, x_{i+1},...,x_n \right), \left(x_1,...,x_{i-1}, x_i^{'}, x_{i+1},...,x_n \right) \in F$ and $t \in [0,1]$, we conclude that F is n-convex set.

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A NEW APPROACH TO THE FRAME DRAGGING EFFECT

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Abstract: *In this paper we consider the effect of frame dragging of inertia. Using an antisymmetric tensor consisting 3-vector of acceleration and 3-vector of angular velocity (analogous to the tensor of electromagnetic field), we derive a formula for frame dragging of inertia, which is the same as the value predicted by the General Relativity. This effect was tested by the Gravity Probe B experiment.*

Key words: *Frame dragging of inertia, Gravity Prove B, General Relativity*

1 INTRODUCTION

In this section we present the basic knowledge about spinning bodies [7, 3]. Let us denote by $\vec{S}$ the spin of a rotating body. It is known that it is Fermi-Walker transported along the world's line. In three dimensions it yields

$$\frac{d\vec{S}}{d\tau} = \vec{\Omega} \times \vec{S}, \qquad (1.1)$$

where

$$\vec{\Omega} = -\frac{1}{2}\frac{\vec{v}\times\vec{a}}{c^2} + (\gamma + \frac{1}{2})\frac{\vec{v}\times\nabla U}{c^2} - \frac{\gamma + 1 + \frac{\alpha_1}{4}}{4} c\nabla \times \vec{g}, \qquad (1.2)$$

and $\vec{g} = g_{4i}\vec{e}_i$. The first term on the right side in (1.2) denotes the Thomas precession and it disappears for free-fall orbit, so it is not of interest now. The second term is the geodetic precession. Notice that the formula for the geodetic precession was recently measured to about 0.7% using Lunar laser ranging data [2, 8] by considering the Earth-Moon system as a gyroscope with its axis perpendicular to the orbital plane.

The third term in (1.2) is the effect of frame dragging of inertia. We can replace there $\gamma = 1$ and $\alpha_1 = 0$ according to the General Relativity. Then [7]

$$\vec{\Omega} = 2\nabla \times \vec{V} \qquad (1.3)$$

where

$$V_i = \frac{G}{c^2}\int \frac{\rho' v_i'}{\mid \vec{x} - \vec{x}' \mid} d^3x'. \qquad (1.4)$$

Solving the corresponding integral, this yields

$$\vec{\Omega} = -\frac{G}{c^2 r^3}[\vec{M} - 3\vec{n}(\vec{M}\vec{n})], \qquad (1.5)$$

where $\vec{M}$ is the angular momentum of the spinning body (the Earth), and $\vec{n}$ is a unit radial vector. In the case of Gravity Probe B, i.e. for polar motion with constant velocity, this vector averages to

$$\langle\vec{\Omega}\rangle = -\frac{G}{2c^2 r^3}\vec{M}. \qquad (1.6)$$

In 1960 Stanford physicist Leonard Schiff (and independently, George Pugh at the Pentagon) proposed an experiment, which would measure the geodetic precession and frame dragging of inertia of gyroscope orbiting around the Earth. Although at that time the experiment seemed to be rather simple, there were needed more than four decades of scientific and technological advance to create a space-borne laboratory and measurement instrument sufficiently sophisticated and precise to measure these two very slight relativistic effects. In April 2004 a spacecraft was launched containing four gyroscopes around the Earth orbiting close the South and North pole on 640 km from the Earth (http://einstein.stanford.edu). The gyroscopes are made in almost ideal spherical form in order to avoid a self motion of the axis of rotation apart from the relativistic effects. The guide star is chosen to be IM Pegasi. The object of experiment is to measure the precession of the gyroscope's spin axis. Two angles of deviation of the spin axis in two orthogonal directions were measured. The first angle measures the geodetic precession, which is expected to be 0.00183 degrees per year, and the second angle measures the frame dragging of inertia, which is expected to be 0.0000114 degrees per year. This test was conducted by the Stanford University, and the results are expected to be known in the middle of April 2007. This test is called Gravity Probe B. However, two unanticipated effects are clouding the GPB team's frame-dragging results [1].

2 DEDUCING THE FORMULA FOR FRAME DRAGGING OF INERTIA IN AN ALTERNATIVE WAY

In this section we present a relativistic approach, close to the Special Relativity, which gives the same value for the frame dragging of inertia as predicted by the General Relativity.

Let us consider one gravitational body with a mass M and a 4-vector of velocity

$$(U_1, U_2, U_3, U_4) = \frac{1}{\sqrt{1-u^2/c^2}}\Big(\frac{u_x}{ic}, \frac{u_y}{ic}, \frac{u_z}{ic}, 1\Big), \qquad (2.1)$$

where $\vec{u} = (u_x, u_y, u_z)$ is a 3-vector of velocity. This leads to the following antisymmetric tensor [6]

$$\phi_{ij} = U_i \frac{1}{\mu}\frac{\partial \mu}{\partial x_j} - U_j \frac{1}{\mu}\frac{\partial \mu}{\partial x_i}, \qquad (2.2)$$

where $\mu = 1 + \frac{GM}{rc^2}$ and r is the distance to the considered test body.

If $(U_i) = (0, 0, 0, 1)$, then $(c^2\phi_{41}, c^2\phi_{42}, c^2\phi_{43})$ represents just the Newtonian acceleration toward the gravitational body, and $\phi_{12} = \phi_{23} = \phi_{31} = 0$. In a general case when $(U_i) \neq (0, 0, 0, 1)$, if $c^2\phi_{41}$, $c^2\phi_{42}$, and $c^2\phi_{43}$ are the angular velocities in the x_1x_4, x_2x_4, x_3x_4-plane respectively, then ϕ_{12}, ϕ_{23}, and ϕ_{31} represent the angular velocities in the xy, yz, and zx-plane respectively. Thus the general structure of (2.2) is given by

$$\phi = \begin{bmatrix} 0 & -i\omega_z/c & i\omega_y/c & -a_x/c^2 \\ i\omega_z/c & 0 & -i\omega_x/c & -a_y/c^2 \\ -i\omega_y/c & i\omega_x/c & 0 & -a_z/c^2 \\ a_x/c^2 & a_y/c^2 & a_z/c^2 & 0 \end{bmatrix}, \qquad (2.3)$$

where $\vec{a} = (a_x, a_y, a_z)$ and $\vec{w} = (w_x, w_y, w_z)$ respectively can be interpreted as a 3-vector of acceleration and a 3-vector of angular velocity. We notice that ϕ is analogous to the electromagnetic tensor, $\vec{a}$ is analogous to the electric 3-vector field $\vec{E}$, and $\vec{w}$ is analogous to the magnetic 3-vector field $\vec{H}$.

Now we will obtain the formula $\vec{\Omega} = -2\vec{w}$ without using the physical meaning of ϕ_{ij}, $(1 \le i, j \le 3)$. We will use the equations of parallel transport along any curve. This (nonlinear) connection is given in [4, 5, 6], and we present it now only for the case of one gravitational body.

Assume that the 4-vector of velocity of a test body with zero mass is given by

$$(V_1, V_2, V_3, V_4) = \frac{1}{\sqrt{1 - v^2/c^2}} \Big(\frac{v_x}{ic}, \frac{v_y}{ic}, \frac{v_z}{ic}, 1 \Big), \qquad (2.4)$$

and let us consider a system of 4 orthonormal vectors A_{i1}, A_{i2}, A_{i3}, and A_{i4}, where $A_{i\alpha}$ is the i-th coordinate of the α-th vector. Indeed, we assumed here that $x_4 = ict$, and an orthogonal matrix with complex coefficients means that $AA^T = I$. Using that $A_{i\alpha}$ is an orthogonal matrix, the following tensor

$$\frac{dA_{i\alpha}}{ds} A_{j\alpha} \qquad (2.5)$$

is also antisymmetric. In (2.5) $ds = ic\sqrt{1 - \frac{v^2}{c^2}}dt$, just like in the Special Relativity. Notice that (2.5) is invariant under the linear transformation $A_{i\alpha} \rightarrow B_{i\alpha} = A_{i\beta} R_{\beta\alpha}$, where R is an orthogonal matrix with constant elements. In the special case when $U_i = V_i$, we assume that the tensors (2.2) and (2.5) are equal. Then the physical interpretation (2.3) of (2.2) is obvious. Since (2.5) is invariant under the linear transformation $A \rightarrow AR$, we can assume that $A_{ij} = \delta_{ij}$ at the considered point, without loss of generality. Hence the components of (2.5) are 3-vector of acceleration and 3-vector of angular velocity.

If $U_i \neq V_i$, then we accept the following relationship between the tensors (2.2) and (2.5)

$$\frac{dA_{i\alpha}}{ds} A_{j\alpha} = P_{ri} \phi_{rk} P_{kj}, \qquad (2.6)$$

or in matrix form $\frac{dA}{ds} A^T = P^T \phi P$. Both sides of (2.6) are antisymmetric matrices, and here $P = P(U, V)$ is a tensor given by

$$P_{ij} = \delta_{ij} - \frac{1}{1 + U_s U_s} (V_i V_j + V_i U_j + U_i V_j + U_i U_j) + 2U_j V_i. \qquad (2.7)$$

This tensor P is an orthogonal matrix and it satisfies $P(V, U) = P^T(U, V)$. Moreover, if $U_i = (0, 0, 0, 1)$, then $P(U, V)$ is just a Lorentz transformation.

Finally, if we multiply the equation (2.6) by $A_{j\beta}$ we get

$$\frac{dA_{i\beta}}{ds} = P_{ri} \phi_{rk} P_{kj} A_{j\beta},$$

and hence for the parallel displacement of arbitrary (unit) vector A_i we get

$$\frac{dA_i}{ds} = P_{ri} \phi_{rk} P_{kj} A_j. \qquad (2.8)$$

Specially, for $A_i = V_i$, we obtain the equations for autoparallel displacement, i.e.

$$\frac{dV_i}{ds} = P_{ri}\phi_{rk}P_{kj}V_j. \qquad (2.9)$$

We will not consider the equations (2.8) and (2.9) in more detail, but we will only mention that these equations are related to the moving coframe, considering the orthonormal tetrades. The calculations [6] show that considering the perihelion shift, deflection of the light ray near the Sun and geodetic precession give the same formulae as the General Relativity. In the rest of the paper we will consider this effect and obtain the same formula for the frame dragging effect as in the General Relativity.

Let us imagine the Earth as a large number of particles (atoms). Let m_i be the mass of the i-th particle and r_i be the distance from its center to the test body, such that r_i is a function of 3 coordinates of the i-th body and 3 coordinates of the considered test body. Further, let the 3-vector of velocity of the i-th particle be $\vec{u}_i = (u_{ix}, u_{iy}, u_{iz})$ and let us put $\mu_i = 1 + \frac{Gm_i}{r_i c^2}$. The index i refers to the i-th particle. Let us denote the 4-vector of velocity of the i-th particle by

$$(U_{i1}, U_{i2}, U_{i3}, U_{i4}) = \Big(\frac{\frac{u_{ix}}{ic}}{\sqrt{1-\frac{u_i^2}{c^2}}}, \frac{\frac{u_{iy}}{ic}}{\sqrt{1-\frac{u_i^2}{c^2}}}, \frac{\frac{u_{iz}}{ic}}{\sqrt{1-\frac{u_i^2}{c^2}}}, \frac{1}{\sqrt{1-\frac{u_i^2}{c^2}}}\Big).$$

The tensor ϕ_{pq} induced by the i-th particle in our coordinate system is given by $\phi_{pq} = \frac{1}{\mu_i}\Big[U_{ip}\frac{\partial \mu_i}{\partial x_q} - U_{iq}\frac{\partial \mu_i}{\partial x_p}\Big]$ and hence

$$w_{ix} = \frac{-ic}{\mu_i}\Big[U_{i3}\frac{\partial \mu_i}{\partial y} - U_{i2}\frac{\partial \mu_i}{\partial z}\Big], \quad w_{iy} = \frac{-ic}{\mu_i}\Big[U_{i1}\frac{\partial \mu_i}{\partial z} - U_{i3}\frac{\partial \mu_i}{\partial x}\Big],$$

$$w_{iz} = \frac{-ic}{\mu_i}\Big[U_{i2}\frac{\partial \mu_i}{\partial x} - U_{i1}\frac{\partial \mu_i}{\partial y}\Big].$$

Thus,

$$\vec{w}_i = -\Big(\frac{u_{i3}}{c^2}\frac{\partial\frac{Gm_i}{r_i}}{\partial y} - \frac{u_{i2}}{c^2}\frac{\partial\frac{Gm_i}{r_i}}{\partial z}, \frac{u_{i1}}{c^2}\frac{\partial\frac{Gm_i}{r_i}}{\partial z} - \frac{u_{i3}}{c^2}\frac{\partial\frac{Gm_i}{r_i}}{\partial x}, \frac{u_{i2}}{c^2}\frac{\partial\frac{Gm_i}{r_i}}{\partial x} - \frac{u_{i1}}{c^2}\frac{\partial\frac{Gm_i}{r_i}}{\partial y}\Big).$$

where the terms of order c^{-4} and smaller are neglected. We assume that the gyroscope has a negligible mass, i.e. 0, we can replace $\vec{v} = 0$ ignoring the geodetic precession. Calculating the antisymmetrc matrix $P^T\phi P$, where

$$P = P(U,0) = P(0,U)^{-1} = \begin{bmatrix} 1-\frac{1}{\nu}U_1^2 & -\frac{1}{\nu}U_1U_2 & -\frac{1}{\nu}U_1U_3 & -U_1 \\ -\frac{1}{\nu}U_2U_1 & 1-\frac{1}{\nu}U_2^2 & -\frac{1}{\nu}U_2U_3 & -U_2 \\ -\frac{1}{\nu}U_3V_1 & -\frac{1}{\nu}U_3U_2 & 1-\frac{1}{\nu}U_3^2 & -U_3 \\ U_1 & U_2 & U_3 & U_4 \end{bmatrix}$$

and $\nu = 1 + U_4$, one easily obtains that the angular velocity of this matrix product is twice larger, i.e. it is $2\vec{w}_i$. After applying the parallel transport of space vectors $(1,0,0,0)$, $(0,1,0,0)$, and $(0,0,1,0)$, it is easy to see that the corresponding angular velocity is $-2\vec{w}_i$, i.e. $\vec{\Omega}_i = -2\vec{w}_i$. Finally, summing up by i for all particles of the Earth we obtain

$$\vec{\Omega} = \sum_i \vec{\Omega}_i =$$

$$-2\sum_i \left(\frac{u_{i3}}{c^2}\frac{\partial \frac{Gm_i}{r_i}}{\partial y} - \frac{u_{i2}}{c^2}\frac{\partial \frac{Gm_i}{r_i}}{\partial z}, \frac{u_{i1}}{c^2}\frac{\partial \frac{Gm_i}{r_i}}{\partial z} - \frac{u_{i3}}{c^2}\frac{\partial \frac{Gm_i}{r_i}}{\partial x}, \frac{u_{i2}}{c^2}\frac{\partial \frac{Gm_i}{r_i}}{\partial x} - \frac{u_{i1}}{c^2}\frac{\partial \frac{Gm_i}{r_i}}{\partial y} \right)$$

$$= 2\sum_i \frac{\vec{a}_i \times \vec{u}_i}{c^2}, \qquad (2.10)$$

where $\vec{a}_i$ is the Newtonian acceleration vector toward the i-th particle. This sum yield an integral which compared with (1.3) where $\vec{V}$ is given by (1.4), is the same value as obtained by the General Relativity.

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GENERALIZED FUZZY CONTINUOUS MAPPINGS

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Abstract*: Some new classes of generalized fuzzy mappings in Chang's fuzzy topological space have been introduced and studied. Their properties and relationships with other early defined classes of generalized fuzzy mappings have been investigated.*

Keywords: *Fuzzy topology, generalized fuzzy regular continuous mapping, generalized fuzzy α-continuous mapping, generalized fuzzy regular α-continuous mapping.*

1. INTRODUCTION

Chang introduced the notion of fuzzy topology in his classical paper [3]. Balasubramanian and Sundaram [2] introduced the concept of fuzzy generalized closed sets in Chang's fuzzy topology as an extension of generalized closed sets of Levine [6] in ordinary topology. More details about the generalized closed sets can be found in [4].

The authors [5] defined the concepts of fuzzy generalized α-closed sets and fuzzy generalized regular α-closed sets in Chang's fuzzy topological space. Here we will define some new classes of generalized fuzzy mappings in Chang's fuzzy topological space. We will investigate their properties and relationships with other early defined classes of generalized fuzzy mappings.

2. PRELIMINARIES

Throughout this paper, by (X,τ) or simply by X will be denoted fuzzy topological space (fts) due to Chang. The reader can be referee to the [1], [4] and [7] for more details about the fuzzy open sets, fuzzy closed sets, fuzzy regular open sets, fuzzy regular closed sets, fuzzy α-open sets and fuzzy α-closed sets. The interior, the closure, the α-interior, the α-closure and the complement of a fuzzy set λ will be denoted by $\operatorname{int}\lambda$, $\operatorname{cl}\lambda$, $\alpha\operatorname{int}\lambda$, $\alpha\operatorname{cl}\lambda$ and $1-\lambda$, respectively.

Definition 2.1. [2,5] Let λ be a fuzzy set of an fts X. Then λ is called

(1) a generalized fuzzy closed set if and only if $\operatorname{cl}\lambda \le \mu$*, for each fuzzy open set* μ *such that* $\lambda \le \mu$;

(2) a generalized fuzzy regular closed set if and only if $\operatorname{cl}\lambda \le \mu$*, for each fuzzy regular open set* μ *such that* $\lambda \le \mu$;

(3) a generalized fuzzy α-closed set if and only if $\alpha\operatorname{cl}\lambda \le \mu$*, for each fuzzy open set* μ *such that* $\lambda \le \mu$;

(4) a generalized fuzzy regular α-closed set if and only if $\alpha\operatorname{cl}\lambda \le \mu$*, for each fuzzy regular open set* μ *such that* $\lambda \le \mu$.

A fuzzy set λ of an fts X is called generalized fuzzy open (generalized fuzzy regular open, generalized fuzzy α-open, generalized fuzzy regular α-open) if and only if $1-\lambda$ is a generalized fuzzy closed (generalized fuzzy regular closed, generalized fuzzy α-closed, generalized fuzzy regular α-closed) set.

Definition 2.2. [2] A mapping $f: X \to Y$ from an fts X into an fts Y is called:

(1) generalized fuzzy continuous if and only if $f^{-1}(\mu)$ is a generalized fuzzy open set in X, for each fuzzy open set μ in Y.

(2) generalized fuzzy open if and only if $f(\lambda)$ is a generalized fuzzy open set in Y, for each fuzzy open set λ in X.

(3) generalized fuzzy closed if and only if $f(\lambda)$ is a generalized fuzzy closed set in Y, for each fuzzy closed set λ in X.

3. Generalized fuzzy continuous mappings

Definition 3.1. A mapping $f: X \to Y$ from an fts X into an fts Y is called:

(1) generalized fuzzy regular continuous (generalized fuzzy α-continuous, generalized fuzzy regular α-continuous) if and only if $f^{-1}(\mu)$ is a generalized fuzzy regular open (generalized fuzzy α-open, generalized fuzzy regular α-open) set in X, for each fuzzy open set μ in Y.

(2) generalized fuzzy regular open (generalized fuzzy α-open, generalized fuzzy regular α-open) if and only if $f(\lambda)$ is a generalized fuzzy regular open (generalized fuzzy α-open, generalized fuzzy regular α-open) set in Y, for each fuzzy open set λ in X.

(3) generalized fuzzy regular closed (generalized fuzzy α-closed, generalized fuzzy regular α-closed) if and only if $f(\lambda)$ is a generalized fuzzy regular closed (generalized fuzzy α-closed, generalized fuzzy regular α-closed) set in Y, for each fuzzy closed set λ in X.

Remark 3.1. *From the above definitions it is not difficult to conclude that the following diagrams of implications is true.*

generalized fuzzy contunuous $\Rightarrow$ *generalized fuzzy regular continuous*
$\Downarrow$ $\Downarrow$
generalized fuzzy α*-continuous* $\Rightarrow$ *generalized fuzzy regular* α*-continuous*

generalized fuzzy open $\Rightarrow$ *generalized fuzzy regular open*
$\Downarrow$ $\Downarrow$
generalized fuzzy α*-open* $\Rightarrow$ *generalized fuzzy regular* α*-open*

generalized fuzzy closed $\Rightarrow$ *generalized fuzzy regular closed*
$\Downarrow$ $\Downarrow$
generalized fuzzy α*-closed* $\Rightarrow$ *generalized fuzzy regular* α*-closed*

The following example shows that the converse may not be true.

Example 3.1. Let $X = \{a, b\}$ and let λ, μ and ν are fuzzy sets defined by $\lambda(a) = 0{,}2$; $\lambda(b) = 0{,}4$; $\mu(a) = 0{,}9$; $\mu(b) = 0{,}4$; $\nu(a) = 0{,}1$; $\nu(b) = 0{,}4$.

Let $\tau_1 = \{0, \lambda, \mu, 1\}$. By easy computation it can be shown that ν is a generalized fuzzy α-closed set, but it is not a generalized fuzzy closed set. Also, ν is a generalized fuzzy regular α-closed set, but it is not a generalized fuzzy regular closed set.

If we put $\tau_2 = \{0, \mu, 1\}$, then the fuzzy set ν is a generalized fuzzy regular α-closed set, but it is not a generalized fuzzy α-closed set. Also, ν is a generalized fuzzy regular closed set, but it is not a generalized fuzzy closed set.

Let $\tau_3 = \{0, 1-\nu, 1\}$. The mapping $id_X : (X, \tau_1) \to (X, \tau_3)$ is generalized fuzzy α-continuous, but it is not generalized fuzzy continuous. Also, $id_X : (X, \tau_1) \to (X, \tau_3)$ is generalized fuzzy regular α-continuous, but it is not generalized fuzzy regular continuous.

The mapping $id_X : (X, \tau_2) \to (X, \tau_3)$ is generalized fuzzy regular α-continuous, but it is not generalized fuzzy α-continuous. Also, $id_X : (X, \tau_2) \to (X, \tau_3)$ is generalized fuzzy regular continuous, but it is not generalized fuzzy continuous.

The mapping $id_X : (X, \tau_3) \to (X, \tau_1)$ is generalized fuzzy α-open (generalized fuzzy α-closed), but it is not generalized fuzzy open (generalized fuzzy closed). Also, $id_X : (X, \tau_3) \to (X, \tau_1)$ is generalized fuzzy regular α-open (generalized fuzzy regular α-closed), but it is not generalized fuzzy regular open (generalized fuzzy regular open).

The mapping $id_X : (X, \tau_3) \to (X, \tau_2)$ is generalized fuzzy regular α-open (generalized fuzzy regular α-closed), but it is not generalized fuzzy α-open (generalized fuzzy α-closed). Also, $id_X : (X, \tau_3) \to (X, \tau_2)$ is generalized fuzzy regular open (generalized fuzzy regular closed), but it is not generalized fuzzy open (generalized fuzzy closed).♦

Theorem 3.1. A mapping $f : X \to Y$ from an fts X into an fts Y is generalized fuzzy regular continuous (generalized fuzzy α-continuous, generalized fuzzy regular α-continuous) if and only if $f^{-1}(\mu)$ is a generalized fuzzy regular closed (generalized fuzzy α-closed, generalized fuzzy regular α-closed) set in X, for each fuzzy closed set μ in Y.

Proof. It can be prove by using the complement. ■

Theorem 3.2. Let $f : X \to Y$ be a bijective mapping from an fts X into an fts Y. Then f is generalized fuzzy regular open (generalized fuzzy α-open, generalized fuzzy regular α-open) if and only if it is generalized fuzzy regular closed (generalized fuzzy α-closed, generalized fuzzy regular α-closed).

Proof. It can be prove by using the complement. ■

Theorem 3.3. Let $f : X \to Y$ be a bijective mapping from an fts X into an fts Y. Then f is generalized fuzzy regular open (generalized fuzzy α-open, generalized fuzzy regular α-open) if and only if f^{-1} is generalized fuzzy regular continuous (generalized fuzzy α-continuous, generalized fuzzy regular α-continuous).

Proof. It follows from the relation $(f^{-1})^{-1}(A) = f(A)$, for each fuzzy open (regular open) set A of X. ■

Corollary 3.4. Let $f : X \to Y$ be a bijective mapping from an fts X into an fts Y. Then f is generalized fuzzy regular closed (generalized fuzzy α-closed, generalized fuzzy regular α-closed) if and only if f^{-1} is generalized fuzzy regular continuous (generalized fuzzy α-continuous, generalized fuzzy regular α-continuous).

Theorem 3.5. Let $f : X \to Y$ be a mapping from an fts X onto an fts Y. Then f is generalized fuzzy regular closed (generalized fuzzy α-closed, generalized fuzzy regular α-closed) if and only if for each fuzzy set ρ in Y and each fuzzy open set μ in X such that $f^{-1}(\rho) \le \mu$, there exists a generalized fuzzy regular open (generalized fuzzy α-open, generalized fuzzy regular α-open) set ν in Y such that $\rho \le \nu$ and $f^{-1}(\nu) \le \mu$.

Proof. Let ρ be any fuzzy set in Y and let μ be a fuzzy open set in X such that $f^{-1}(\rho) \le \mu$. We put $\nu = 1 - f(1-\mu)$. Then $f(1-\mu)$ is a generalized fuzzy regular closed

set, so ν is a generalized fuzzy regular open set. From $f^{-1}(\rho) \le \mu$ follows that $1-\mu \le 1-f^{-1}(\rho) = f^{-1}(1-\rho)$. Hence
$f(1-\mu) \le f(1-f^{-1}(\rho)) \le 1-\rho$. Therefore $\rho \le 1-f(1-\mu) = \nu$ and
$f^{-1}(\nu) = f^{-1}(1-f(1-\mu) = 1-f^{-1}(f(1-\mu)) \le 1-(1-\mu) = \mu.$

Conversely, let μ be any fuzzy closed set. Then $1-\mu$ is a fuzzy open set and $f^{-1}(1-f(\mu)) \le 1-f^{-1}(f(\mu)) \le 1-\mu$. Hence there exists a generalized fuzzy regular open set ν in X such that $1-f(\mu) \le \nu$ and $f^{-1}(\nu) \le 1-\mu$. It follows that $\mu \le 1-f^{-1}(\nu) = f^{-1}(1-\nu)$. Hence $f(\mu) \le f(1-f^{-1}(\nu)) \le 1-\nu$. Thus $f(\mu) = 1-\nu$ is a generalized fuzzy regular closed set, so f is a generalized fuzzy regular closed mapping.

The other cases can be proved in a similar manner. ■

Theorem 3.6. Let $f: X \to Y$ be a mapping from an fts X onto an fts Y. Then f is generalized fuzzy regular open mapping (generalized fuzzy α-open, generalized fuzzy regular α-open) if and only if for each fuzzy set ρ in Y and each fuzzy closed set μ in X such that $f^{-1}(\rho) \le \mu$, there exists a generalized fuzzy regular closed (generalized fuzzy α-closed, generalized fuzzy regular α-closed) set ν in Y such that $\rho \le \nu$ and $f^{-1}(\nu) \le \mu$.

Proof. It can be proved in a similar manner as the Theorem 3.5. ■

Corollary 3.7. Let $f: X \to Y$ be a mapping from an fts X onto an fts Y. If f is a generalized fuzzy regular closed (generalized fuzzy α-closed, generalized fuzzy regular α-closed) mapping then for each fuzzy regular closed set ρ in Y and each fuzzy regular open set μ in X such that $f^{-1}(\rho) \le \mu$, there exists a fuzzy open (fuzzy α-open) set ν in Y such that $\rho \le \nu$ and $f^{-1}(\nu) \le \mu$.

Proof. Let f be a generalized fuzzy α-closed mapping. Suppose ρ be any fuzzy regular closed set and let μ be a fuzzy regular open set such that $f^{-1}(\rho) \le \mu$. From the Theorem 3.5. follows that there exists generalized fuzzy α-open set γ, such that $\rho \le \gamma$ and $f^{-1}(\gamma) \le \mu$. Since γ is a generalized fuzzy α-open set and $\rho \le \gamma$ we obtain that $\rho \le \alpha \operatorname{int} \gamma$. Then $\nu = \alpha \operatorname{int} \gamma$ is a fuzzy α-open set such that $\rho \le \nu$ and $f^{-1}(\nu) \le \mu$.

The other cases can be proved in a similar manner. ■

Theorem 3.8. Let $f: X \to Y$ and $g: Y \to Z$ be mappings where X, Y and Z are fts's. If g is a fuzzy continuous mapping and f is a generalized fuzzy regular continuous (generalized fuzzy α-continuous, generalized fuzzy regular α-continuous) mapping, then gf is a generalized fuzzy regular continuous (generalized fuzzy α-continuous, generalized fuzzy regular α-continuous) mapping.

Proof. From $(gf)^{-1}(\mu) = f^{-1}(g^{-1}(\mu))$, for any open fuzzy μ in Z follows that $(gf)^{-1}(\mu)$ is a generalized fuzzy regular set in X.

The other cases can be proved in a similar manner. ■

Theorem 3.9. Let $f: X \to Y$ and $g: Y \to Z$ be mappings where X, Y and Z are fts's. If g is a generalized fuzzy regular open (generalized fuzzy α-open, generalized fuzzy regular α-open) mapping and f is fuzzy open mapping then gf is a generalized fuzzy regular open (a generalized fuzzy α-open, a generalized fuzzy regular α-open) mapping.

Proof. For any fuzzy open set μ in X holds $(gf)(\mu) = g(f(\mu))$. Since f is a fuzzy open mapping and g is a generalized fuzzy regular open mapping we obtain that $g(f(\mu))$ is a generalized fuzzy regular set in Z.

The other cases can be proved in a similar manner. ■

Theorem 3.10. Let $f : X \to Y$ and $g : Y \to Z$ be mappings where X, Y and Z are fts's. If g is a generalized fuzzy regular closed (generalized fuzzy α-closed, generalized fuzzy regular α-closed) mapping and f is a fuzzy closed mapping then gf is a generalized fuzzy regular closed (generalized fuzzy α-closed, generalized fuzzy regular α-closed) mapping.

Proof. It can be proved in a similar manner as the Theorem 3.9. ■

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FREE OBJECTS IN THE VARIETY OF GROUPOIDS DEFINED BY THE IDENTITY $xx^{(m)} \approx x^{(m+1)}$

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Abstract*: A construction of free objects in the variety $\mathcal{V}_{(m)}$ of groupoids defined by the identity $xx^{(m)} \approx x^{(m+1)}$, where m is a fixed positive integer, and (k) is a transformation of a groupoid $\boldsymbol{G} = (G, \cdot)$ defined by $x^{(0)} = x$, $x^{(k+1)} = (x^{(k)})^2$, is given. A class of injective groupoids in $\mathcal{V}_{(m)}$ is defined and a corresponding Bruck theorem for this variety is proved. It is shown that the class of free groupoids in $\mathcal{V}_{(m)}$ is a proper subclass of the class of injective groupoids in $\mathcal{V}_{(m)}$.*

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1 PRELIMINARIES

Let $\boldsymbol{G} = (G, \cdot)$ be a groupoid, i.e. an algebra with one binary operation.

For any nonnegative integer k we define a transformation $(k) : x \mapsto x^{(k)}$ of G as follows:

$$x^{(0)} = x, \qquad x^{(k+1)} = (x^{(k)})^2.$$

(Here, x^k is defined by: $x^1 = x$, $x^{k+1} = x^k x$; ex: $x^4 = ((xx)x)x$.)

We say that $x^{(k)}$ is the k *square of* x and $x^{(1)} = x^2$ the *square* of x.

By induction on p and q one can show ([2]) that in any groupoid $\boldsymbol{G} = (G, \cdot)$ $(x^{(p)})^{(q)} = x^{(p+q)}$ for any $x \in G$ and any nonnegative integers p, q.

The variety of groupoids defined by the identity $xx^{(m)} \approx x^{(m+1)}$ will be denoted by $\mathcal{V}_{(m)}$, for a fixed positive integer m. (The variety $\mathcal{V}_{(1)}$ is investigated in [3].)

If $\boldsymbol{G} \in \mathcal{V}_{(m)}$, then by induction on p, one obtains:

$$(\forall x \in G, p \geq 0) \;\; x^{(p)} x^{(p+m)} = x^{(p+m+1)}.$$

An element $a \in G$ is said to be *2-primitive* in $\boldsymbol{G}$ if and only if

$$(\forall x \in G, p \geq 0) \;\; (a = x^{(p)} \;\Rightarrow\; p = 0).$$

In the sequel B will be an arbitrary nonempty set whose elements are called variables. By T_B we will denote the set of all groupoid terms over B in the signature $\cdot$. The terms are denoted by $t, u, v, \ldots, x, y, \ldots$ $\boldsymbol{T}_B = (T_B, \cdot)$ is the absolutely free groupoid with the free basis B, where the operation is defined by $(u, v) \mapsto uv$. It is well known (Bruck theorem for $\boldsymbol{T}_B$, [1]) that the following two properties characterize $\boldsymbol{T}_B$:

(i) $\boldsymbol{T}_B$ is *injective*, i.e. the operation $\cdot : (u, v) \mapsto uv$ is an injection.

(ii) The set B of primes in $\boldsymbol{T}_B$ is nonempty and generates $\boldsymbol{T}_B$.

(An element a in a groupoid $\boldsymbol{G} = (G, \cdot)$ is said to be *prime* in $\boldsymbol{G}$ if and only if $a \neq xy$, for any $x, y \in G$.)

For any term v of T_B we define the *length* $|\, v\,|$ of v and the *set of subterms* $P(v)$ of v in the following way:

$$|\, b\,| = 1,\ |\, tu\,| = |\, t\,| + |\, u\,|;\ P(b) = \{b\},\ P(tu) = \{tu\} \cup P(t) \cup P(u),$$

for any variable b and any terms t, u of T_B.

Bellow we consider a few properties of $x^{(k)}$ in $\boldsymbol{T}_B$ that can be shown by induction on p.

Proposition 1.1 *If* $t, u \in T_B$ *and* p, q *are nonnegative integers, then:*

a) $|\, t^{(p)}\,| = 2^p |\, t\,|$.

b) $t^{(p)} = u^{(p+q)} \Rightarrow t = u^{(q)}$.

c) If t, u *are 2-primitive elements in* T_B*, then:* $t^{(p)} = u^{(q)} \Leftrightarrow t = u,\ p = q$.

Proof. c) We assume that $p \leq q$, i.e. $q = p + k$, for any $k \geq 0$. Then from $t^{(p)} = u^{(q)}$ we have that $t^{(p)} = u^{(p+k)}$. By b) it follows that $t = u^{(k)}$. However, t is 2-primitive in $\boldsymbol{T}_B$, therefore $k = 0$. Thus $t = u^{(0)} = u$, and $p = q$. The converse is obvious. □

By Prop.1 c) it follows directly that:

Proposition 1.2 *For any* $t \in T_B$*, there is a unique 2-primitive term* $\alpha \in T_B$ *and a unique nonnegative integer* p*, such that* $t = \alpha^{(p)}$. □

We say that α is the 2-*base* of t and p is the 2-*exponent* of t; we denote them by $\underline{t} = \alpha$, $[t] = p$, respectively.

2 A construction of free objects in $\mathcal{V}_{(m)}$

Assuming that B is a nonempty set and $\boldsymbol{T}_B = (T_B, \cdot)$ the absolutely free groupoid with the free basis B, we are looking for a *canonical groupoid* ([4]) in $\mathcal{V}_{(m)}$, i.e. a groupoid $\boldsymbol{R} = (R, *)$ with the following properties:

$i)$ $B \subset R \subset T_B$; $ii)$ $tu \in R \Rightarrow t, u \in R$; $iii)$ $tu \in R \Rightarrow t * u = tu$

$iv)$ $\boldsymbol{R}$ is a free groupoid in $\mathcal{V}_{(m)}$ with the free basis B.

Define the carrier R of the desired groupoid $\boldsymbol{R}$ by:

$$R = \{t \in T_B : \ (\forall x \in T_B) \ \ xx^{(m)} \notin P(t)\}. \qquad (1)$$

The following properties of R are obvious corollaries of (2.1).

Proposition 2.1 a) R satisfies $i)$ and $ii)$.

b) $t, u \in R \ \Rightarrow \ \{tu \notin R \Leftrightarrow \ u = t^{(m)}\}$.

c) $t, u \in T_B \ \Rightarrow \ \{tu \in R \Leftrightarrow t, u \in R \ \ \& \ \ u \neq t^{(m)}\}$.

d) $t \in R, \ p \geq 1 \ \Rightarrow \ t^p \in R, \ t^{(p)} \in R$. ☐

We define an operation $*$ on R as follows.

$$t, u \in R \ \Rightarrow t * u = \begin{cases} tu, & \text{if } \ tu \in R \\ t^{(m+1)}, & \text{if } \ u = t^{(m)}. \end{cases} \qquad (2)$$

From (2.2) and Prop.2.1 d), by induction on p, we obtain:

e) $t \in R, \ p \geq 1 \ \Rightarrow \ t_*^p = t^p, \ \ t_*^{(p)} = t^{(p)}$,

where t_*^k is defined by: $t_* = t$, $t_*^{k+1} = t_*^k * t$ and $t_*^{(p)}$ is the p square of t in $\boldsymbol{R}$.

By a direct verification one can show that the operation $*$ is well-defined, i.e. $\boldsymbol{R} = (R, *)$ is a groupoid. From (2.2) it follows that if $tu \in R$, then $t, u \in R \ \ \& \ \ t * u = tu$ (i.e. $\boldsymbol{R}$ satisfies $ii)$ and $iii)$). By the property e) and (2.2), we obtain that $t * t_*^{(m)} = t * t^{(m)} = t^{(m+1)} = t_*^{(m+1)}$, i.e. $\boldsymbol{R} \in \mathcal{V}_{(m)}$.

The set of primes in $\boldsymbol{R}$ coincides with B and generates $\boldsymbol{R}$. Namely, every $b \in B$ is prime in $\boldsymbol{R}$, since $b \neq t * u$, for any $t, u \in R$. To show that no element of $R \setminus B$ is prime in $\boldsymbol{R}$, let $t \in T_B \setminus B$ be a term belonging to R. Then there are $t_1, t_2 \in T_B$, such that $t = t_1 t_2$. By the fact that $t \in R$, i.e. $t_1 t_2 \in R$, it follows that $t_1, t_2 \in R$ and $t = t_1 t_2 = t_1 * t_2$, i.e. t is not prime in $\boldsymbol{R}$. Let $\boldsymbol{Q}$ be the subgroupoid of $\boldsymbol{R}$ generated by B, $\boldsymbol{Q} = \langle B \rangle_*$. We will show that $R = Q$. Clearly, $Q \subseteq R$. To show that $R \subseteq Q$, let $t \in R$. If $t \in B$, then $t \in \langle B \rangle_* = Q$, i.e. $(t \in R \ \& \ |\, t \,| = 1 \ \Rightarrow \ t \in Q)$. Suppose that $(t \in R \ \& \ |\, t \,| \leq k \ \Rightarrow \ t \in Q)$ is true. If $t \in R$ is such that $|\, t \,| = k + 1$, then $t = t_1 t_2$ in $\boldsymbol{T}_B$ and $|\, t_1 \,|, |\, t_2 \,| \leq k$. By the inductive hypothesis we have $t_1, t_2 \in Q$, and since $\boldsymbol{Q}$ is a groupoid, it follows that $t = t_1 t_2 = t_1 * t_2 \in Q$. Thus, $R \subseteq Q$. Therefore, $\boldsymbol{R} = \boldsymbol{Q} = \langle B \rangle_*$.

$\boldsymbol{R}$ has the universal mapping property ([5]) for $\mathcal{V}_{(m)}$ over B. Namely, let $\boldsymbol{G} \in \mathcal{V}_{(m)}$, $\lambda : B \to G$ be any mapping and $\varphi : T_B \to G$ be the homomorphism from $\boldsymbol{T}_B$ into $\boldsymbol{G}$ that extends λ. Let $t, u \in R$. If $tu \in R$, then $\varphi(t * u) = \varphi(tu) = \varphi(t)\varphi(u)$. If $tu \notin R$, then $u = t^{(m)}$ and $t * u = t^{(m+1)}$. Using the fact that $\varphi(t^{(p)}) = (\varphi(t))^{(p)}$ (it can be shown by induction on p), we obtain that $\varphi(t*u) = \varphi(t^{(m+1)}) = \varphi(t^{(m)}t^{(m)}) = \varphi(t^{(m)})\varphi(t^{(m)}) = (\varphi(t))^{(m)}(\varphi(t))^{(m)} = (\varphi(t))^{(m+1)} = \varphi(t^{(m+1)}) = [G \in \mathcal{V}_{(m)}] = \varphi(tt^{(m)}) = \varphi(tu) = \varphi(t)\varphi(u)$.

Thus, $\varphi|_R : R \to G$ is a homomorphism that extends λ.

Therefore, the conditions $i)$ - $iv)$ at the begining of this section are fulfilled and thus we proved the following

Theorem 2.1 *The groupoid* $\boldsymbol{R} = (R, *)$, *defined by* (2.1) *and* (2.2), *is a canonical groupoid in* $\mathcal{V}_{(m)}$ *with a free basis* B. □

As a consequence of the property e) and the definition of 2-primitive element we obtain the following

Proposition 2.2 *For any* $u \in R$, *there are a unique 2-primitive element* $t \in R$ *and a unique positive integer* p, *such that* $u = t_*^{(p)} = t^{(p)}$. □

By a direct verification one can show that $(R, *)$ is a left cancellative groupoid. $(R, *)$ is not a right cancellative gropoid (ex: $t^{(1)} * t^{(m+1)} = t^{(m+2)} = t^{(m+1)} * t^{(m+1)}$; however $t^{(1)} \neq t^{(m+1)}$).

The following proposition will be used in the next section.

Proposition 2.3 *Let* $x \in R \setminus B$.

a) If x *is a 2-primitive element in* $\boldsymbol{R}$ *or* $x = \alpha^{(p)}$, *where* $[\alpha] = 0$, $1 \leq p \leq m$, *then there is a unique pair* $(u, v) \in R \times R$, *such that* $x = u * v$. (In that case $x = uv$ and $v \neq u^{(m)}$.)

We say that (u, v) is *the pair of divisors* of x in $\boldsymbol{R}$.

b) If $x = t^{(m+p+1)}$, $p \geq 0$, *then* $x = t^{(p+m)} * t^{(p+m)} = t^{(p)} * t^{(p+m)}$.

Thus $(t^{(p+m)}, t^{(p+m)})$ and $(t^{(p)}, t^{(p+m)})$ are pairs of divisors of x. □

3 *Injective objects in* $\mathcal{V}_{(m)}$

In this Section we will give a characterization of the free groupoids in $\mathcal{V}_{(m)}$ by a wider class, called the class of injective groupoids ([4]) in $\mathcal{V}_{(m)}$. For that purpose, we use the properties of the corresponding canonical groupoid $(R, *)$ in $\mathcal{V}_{(m)}$

previously constructed, that concern the non-prime elements in $\boldsymbol{R}$, i.e. elements of $R \setminus B$.

We say that a groupoid $\boldsymbol{H} = (H, \cdot)$ is *injective* in $\mathcal{V}_{(m)}$ (i.e. $\mathcal{V}_{(m)}$-*injective*) if and only if the following conditions are satisfied:

(0) $\boldsymbol{H} \in \mathcal{V}_{(m)}$

(1) For any $a \in H$, there is a unique 2-primitive element $c \in H$ and a unique nonnegative integer k, such that $a = c^{(k)}$.

(We say that c is the 2-*base* of a and $k = [a]$ is the 2-*exponent* of a.)

(2) If $a \in H$ is a non-prime 2-primitive element in $\boldsymbol{H}$, then there is a unique pair $(c, d) \in H \times H$ such that $a = cd$ and ($\underline{c} \neq \underline{d} \ \vee \ (\underline{d} = \underline{c} \ \& \ [d] \neq m)$).

(In that case we say that (c, d) is the *pair of divisors* of a (we write $(c, d)|a$).)

(3) If $a \in H$ is such that $a = c^{(1)}$, $[c] = p$, $p \leq m - 1$, then (c, c) is the pair of divisors of a.

(4) $a^{(p+m+1)} = cd \ \& \ p \geq 0 \Leftrightarrow [c = d = a^{(p+m)} \vee (c = a^{(p)} \ \& \ d = a^{(p+m)})]$.

From the definition of $\mathcal{V}_{(m)}$-injective groupoid and Prop. we obtain that

Proposition 3.1 *The class of free groupoids in* $\mathcal{V}_{(m)}$ *is a subclass of the class of* $\mathcal{V}_{(m)}$*-injective groupoids.* □

Theorem 3.1 (Bruck Theorem for $V_(m)$) *A groupoid* $\boldsymbol{H}$ *is free in* $\mathcal{V}_{(m)}$ *if and only if* $\boldsymbol{H}$ *satisfies the following two conditions:*

(i) $\boldsymbol{H}$ *is* $\mathcal{V}_{(m)}$*-injective*

(ii) The set P *of primes in* $\boldsymbol{H}$ *is nonempty and generates* $\boldsymbol{H}$.

Proof. If $\boldsymbol{H}$ is free in $\mathcal{V}_{(m)}$ with a free basis B, then by Prop.3, $\boldsymbol{H}$ is $\mathcal{V}_{(m)}$-injective, and by the proof of Theorem , B is the set of primes in $\boldsymbol{H}$ and generates $\boldsymbol{H}$.

For the converse, it suffices to show that $\boldsymbol{H}$ has the universal mapping property for $\mathcal{V}_{(m)}$ over P. Therefore, define an infinite sequence of subsets $C_0, C_1, \ldots$ of H by:

$C_0 = P$, $C_1 = C_0 C_0 = PP$,

$C_{k+1} = \{t \in H \setminus P : \ (c, d)|t \ \Rightarrow \{c, d\} \subseteq C_0 \cup C_1 \cup \cdots \cup C_k \ \& \ \{c, d\} \cap C_k \neq \varnothing\}$.

Then the following statements are true ([4]):

1) $(\forall k \geq 0) \ C_k \neq \varnothing$; 2) $a \in C_k \Rightarrow (\forall p \in \mathbb{N}) \ a^{(p)} \in C_{k+p}, \ \ k \geq 0$.

3) $p \neq q \Rightarrow C_p \cap C_q = \varnothing$; 4) $H = \bigcup\{C_k : k \geq 0\}$.

Let $\boldsymbol{G} \in \mathcal{V}_{(m)}$ and $\lambda : P \to G$ be a mapping. For any nonnegative integer k define a mapping $\varphi_k : C_k \to G$ by $\varphi_0 = \lambda$, and let φ_i be defined for each $i \leq k$. Let $a \in C_{k+1}$ and $(c,d)|a$ are such that $c \in C_r$, $d \in C_s$. Then $r, s \leq k$. If we put $\varphi_{k+1}(a) = \varphi_r(c)\varphi_s(d)$, then $\varphi = \bigcup\{\varphi_i : i \geq 0\}$ is a well defined mapping from H into G. Also, by induction on k we have: $\varphi(a^k) = (\varphi(a))^k$ and $\varphi(a^{(k)}) = (\varphi(a))^{(k)}$, for each $a \in H$ and $k \geq 0$.

If $a \in H$ is a 2-primitive element of $\boldsymbol{H}$ and $(c,d)|a$, then $\varphi(a) = \varphi(c)\varphi(d)$.

If $a \in H$ is such that $a = c^{(1)}$, $[c] = p$, $p \leq m-1$, then $\varphi(a) = \varphi(cc) = \varphi(c)\varphi(c)$.

If $c, d \in H$ are such that $c = d = a^{(p+m)}$, where $p \geq 0$, $a \in H$, then:

$$\varphi(cd) = \varphi(a^{(p+m+1)}) = \varphi((a^{(p+m)})^{(1)}) = (\varphi(a^{(p+m)}))^{(1)} = \varphi(c)\varphi(d).$$

If $c, d \in H$ are such that $c = a^{(p)}$, $d = a^{(p+m)}$, where $p \geq 0$, $a \in H$, then:

$$\varphi(cd) = \varphi(a^{(p+m+1)}) = \varphi((a^{(p+m)})^{(1)}) = \varphi(a^{(p+m)})^{(1)} = \varphi(a)^{(p+m)}\varphi(a)^{(p+m)} =$$
$$(\varphi(a)^{(p)})^{(m)}(\varphi(a)^{(p)})^{(m)} = [\boldsymbol{G} \in \mathcal{V}_{(m)}] = (\varphi(a))^{(p)}(\varphi(a))^{(p+m)} =$$
$$= \varphi(a^{(p)})\varphi(a^{(p+m)}) = \varphi(c)\varphi(d).$$

Thus, in all possible cases we have $\varphi(cd) = \varphi(c)\varphi(d)$, i.e. φ is a homomorphism from $\boldsymbol{H}$ into $\boldsymbol{G}$. Therefore, $\boldsymbol{H}$ is a free groupoid in $\mathcal{V}_{(m)}$ with a free basis P. $\square$

We will give an example of a $\mathcal{V}_{(m)}$-injective groupoid that is not free in $\mathcal{V}_{(m)}$. Let A be an infinite set and $H = A \times \mathbb{N}_0$ ($\mathbb{N}_0$ is the set of nonnegative integers). We will denote the elements of H by a_n instead of (a,n). Define a partial operation $\bullet$ on H by:

(i) $a_p \bullet a_p = a_{p+1}$, (ii) $a_p \bullet a_{p+m} = a_{p+m+1}$,

for any $p \geq 0$ and a fixed positive integer m.

Define a set $D \subseteq H \times H$ by:

$$D = \{(a_k, b_n) : a, b \in A \ \& \ k, n \in \mathbb{N}_0 \ \& \ (a \neq b \ \vee \ (a = b \ \& \ k \neq p \ \& \ n \neq p \ \& \ n \neq p+m, p \geq 0))\}$$

Since $D \sim A \times \{0\}$, there is an injection $\varphi : D \to A \times \{0\}$ and we can put

(iii) $(\forall (a_k, b_n) \in D) \ a_k \bullet b_n = (\varphi(a_k, b_n))_0$.

By a direct verification we obtain that $(H, \bullet)$ is $\mathcal{V}_{(m)}$-injective groupoid. If φ is a bijection, then the set of primes in $\boldsymbol{H}$, i.e. $A \times \{0\} \setminus im\varphi$, is empty. Therefore, by the Bruck Theorem for $\mathcal{V}_{(m)}$, it follows that $(H, \bullet)$ is not free in $\mathcal{V}_{(m)}$. This and Prop.3 proves the following

Proposition 3.2 *The class of free groupoids in $\mathcal{V}_{(m)}$ is a proper subclass of the class of $\mathcal{V}_{(m)}$-injective groupoids.* ☐

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FREE $(m+k,m)-$RECTANGULAR BANDS WHEN $k<m$

Valentina Miovska, Donco Dimovski

Abstract: *A characterization of* $(m+k,m)-$ *rectangular bands when* $k<m$*, using the usual rectangular bands is given in [4]. This result is used to obtain a free* $(m+k,m)-$ *rectangular band when* $k<m$.

Keywords: *rectangular band,* $(m+k,m)-$ *rectangular band, free* $(m+k,m)-$ *rectangular band*

1. INTRODUCTION

First, we will introduce some notations which will be used further on:

1) The elements of Q^s, where Q^s denotes the $s-$th Cartesian power of Q, will be denoted by x_1^s.
2) The symbol x_i^j will denote the sequence $x_i x_{i+1}...x_j$ when $i \le j$, and the empty sequence when $i > j$.
3) If $x_1 = x_2 = ... = x_s = x$, then x_1^s is denoted by the symbol $\overset{s}{x}$.
4) The set $\{1,2,...,s\}$ will be denoted by $\mathbf{N_s}$.

Let $Q \ne \varnothing$ and n,m be positive integers. If [ ] is a map from Q^n into Q^m, then [ ] is called an $(n,m)-$operation. A pair $(Q;[\ ])$ where [ ] is an $(n,m)-$operation is said to be an $(n,m)-$groupoid. Every $(n,m)-$operation on Q induces a sequence $[\ ]_1,[\ ]_2,...,[\ ]_m$ of $n-$ary operations on the set Q, such that

$$((\forall i \in \mathbf{N_m})\ \lfloor x_1^n \rfloor_i = y_i) \Leftrightarrow \lfloor x_1^n \rfloor = y_1^m .$$

Let $m \ge 2, k \ge 1$. An $(m+k,m)-$groupoid $(Q;[\ ])$ is called an $(m+k,m)-$semigroup if for each $i \in \{0,1,2,...,k\}$

$$\left[x_1^i \left[x_{i+1}^{i+m+k}\right] x_{i+m+k+1}^{m+2k}\right] = \left[\left[x_1^{m+k}\right] x_{m+k+1}^{m+2k}\right].$$

Let $(A;[\ ])$ be an $(m+k,m)-$groupoid, where [ ] is an $(m+k,m)-$operation defined by $[x_1^{m+k}] = x_1^m$. Then $(A;[\ ])$ is an $(m+k,m)-$semigroup and it is called a left-zero $(m+k,m)-$semigroup. Dually, a right-zero $(m+k,m)-$semigroup $(B;[\ ])$ is defined by the operation $[x_1^{m+k}] = x_{k+1}^{m+k}$.

The pair $(A \times B;[\ ])$, where [ ] is an $(m+k,m)-$operation on $A \times B$ defined by

$$[x_1^{m+k}] = y_1^m \ \Leftrightarrow\ (x_i = (a_i,b_i), y_j = (a_j,b_{j+k}), i \in \mathbf{N_{m+k}},\ j \in \mathbf{N_m})$$

is an $(m+k,m)-$semigroup and it is a direct product of a left-zero and a right-zero $(m+k,m)-$semigroup on A and B, respectively. Such an $(m+k,m)-$semigroup is called $(m+k,m)-$rectangular band.

The following propositions characterizes $(m+k,m)-$rectangular bands when $k<m$.

Propositon 1.1 ([4, Proposition 2]) Let $\mathbf{Q}=(Q;[\ ])$ be an $(m+k,m)-$semigroup, $k<m$. $\mathbf{Q}$ is an $(m+k,m)-$rectangular band if and only if the conditions

(a) $\left[x_1^{m+2k}\right]_i=\left[x_1^i x_{i+k+1}^{m+2k}\right]_i$, $i\in \mathbf{N_m}$

(b) $\left[x_1^{m+k}\right]_i=\left[y_1^{j-1}x_i y_{j+1}^{j+k-1}x_{i+k}y_{j+k+1}^{m+k}\right]_j$, $i,j\in \mathbf{N_m}$

(c) $\left[\overset{m+k}{x}\right]=\overset{m}{x}$

are satisfied in $\mathbf{Q}$.

Propositon 1.2 ([4, Proposition 3]) Let $\mathbf{Q}=(Q;[\ ])$ be an $(m+k,m)-$semigroup, $k<m$. Then $\mathbf{Q}$ is a direct product of a left-zero and a right-zero $(m+k,m)-$semigroup if and only if there is a rectangular band $(Q;*)$, such that $\left[x_1^{m+k}\right]_i=x_i*x_{i+k}$, $x_1^{m+k}\in Q^{m+k}$, $i\in \mathbf{N_m}$.

Propositon 1.2 gives a characterization of $(m+k,m)-$rectangular bands using the usual rectangular bands. Rectangular band is a semigroup which is a direct product of a left-zero and a right-zero semigroup, or equivalent, rectangular band is a semigroup $(Q;*)$ that satisfies the following two identities $x*y*z=x*z$ and $x*x=x$, for each $x,y,z\in Q$.

This result of Propositon 1.2 is used to obtain a free $(m+k,m)-$rectangular band when $k<m$.

2. Free $(m+k,m)-$rectangular bands when $k<m$

Let $(Q;*)$ be a free rectangular band with a basis B. Then $Q=B\cup\{ab\,|\,a,b\in B,a\neq b\}$ and operation $*$ is defined by:

$$x * y = \begin{cases} a & \begin{matrix} x = y = a \\ x = a, y = ca \\ x = ac, y = a \\ x = ac, y = da \end{matrix} \\ ab & \begin{matrix} x = a, y = b \\ x = a, y = cb \\ x = ac, y = b \\ x = ac, y = db \end{matrix} \quad , a \neq b \end{cases}$$

Let $k < m$ and let $[\ ]$ be the $(m+k, m)-$operation on Q defined by

$$\left[x_1^{m+k}\right]_i = x_i * x_{i+k}\ ,\ x_1^{m+k} \in Q^{m+k},\ i \in \mathbf{N_m}.$$

Then

Proposition 2 $(Q;[\ ])$ is a free $(m+k, m)-$rectangular band when $k < m$ with a basis B.

Proof. Since $k < m$, let $k + t = m$, $t \geq 1$.

First, we will prove that $\left[\left[x_1^{m+k}\right]x_{m+k+1}^{m+2k}\right]_i = x_i * x_{i+2k}$.

a) Let $i \leq t$. Then $i + k \leq t + k = m$.

We have

$\left[\left[x_1^{m+k}\right]x_{m+k+1}^{m+2k}\right]_i =$

$= \left[x_1^{m+k}\right]_i * \left[x_1^{m+k}\right]_{i+k} =$

$= \left(x_i * x_{i+k}\right) * \left(x_{i+k} * x_{i+2k}\right) =$

$= x_i * x_{i+2k}$

b) Let $t < i \leq m$. Then $i = t + \lambda$, $1 \leq \lambda \leq k$ and $i + k = t + \lambda + k = m + \lambda$.

$\left[\left[x_1^{m+k}\right]x_{m+k+1}^{m+2k}\right]_i =$

$= \left[x_1^{m+k}\right]_i * x_{m+k+\lambda} =$

$= \left(x_i * x_{i+k}\right) * x_{m+k+\lambda} =$

$= x_i * x_{m+k+\lambda} =$

$= x_i * x_{i+2k}$.

Further on we will prove that $\left[x_1^j\left[x_{j+1}^{j+m+k}\right]x_{j+m+k+1}^{m+2k}\right]_i = x_i * x_{i+2k}$.

c) Let $i \leq j$. Then $i \leq j < j + t$ implies $i + k < j + t + k = j + m$. Moreover, $i + k > k \geq j$ i.e. $j < i + k < j + m$. Let $i + k = j + \lambda$.

We obtain

$\left[x_1^j\left[x_{j+1}^{j+m+k}\right]x_{j+m+k+1}^{m+2k}\right]_i =$

$= \left[x_1^i x_{i+1}^j \left[x_{j+1}^{j+m+k}\right]_1 \ldots \left[x_{j+1}^{j+m+k}\right]_\lambda \left[x_{j+1}^{j+m+k}\right]_{\lambda+1} \ldots \left[x_{j+1}^{j+m+k}\right]_m x_{j+m+k+1}^{m+2k}\right]_i =$

$= x_i * \left[x_{j+1}^{j+m+k}\right]_\lambda =$

$= x_i * \left(x_{j+\lambda} * x_{j+\lambda+k}\right) =$

$= x_i * x_{j+\lambda+k} =$

$= x_i * x_{i+2k}$.

d) Let $j < i$.

d1) Let $j+1 \le i \le j+t$ i.e. $i = j+\lambda$, $1 \le \lambda \le t$.

Then $i+k = j+\lambda+k \le j+t+k = j+m$.

$$\left[x_1^j \left[x_{j+1}^{j+m+k}\right] x_{j+m+k+1}^{m+2k}\right]_i =$$
$$= \left[x_{j+1}^{j+m+k}\right]_\lambda * \left[x_{j+1}^{j+m+k}\right]_{\lambda+k} =$$
$$= \left(x_{j+\lambda} * x_{j+\lambda+k}\right) * \left(x_{j+\lambda+k} * x_{j+\lambda+k+k}\right) =$$
$$= x_{j+\lambda} * x_{j+\lambda+k+k} =$$
$$= x_i * x_{i+2k}.$$

d2) Let $j+t < i$ i.e. $i = j+t+\lambda$, $1 \le \lambda \le k-j$. Then $j+t+k < i+k$ i.e. $j+m < i+k$.

$$\left[x_1^j \left[x_{j+1}^{j+m+k}\right] x_{j+m+k+1}^{m+2k}\right]_i =$$
$$= \left[x_{j+1}^{j+m+k}\right]_{t+\lambda} * x_{j+m+k+\lambda} =$$
$$= \left(x_{j+t+\lambda} * x_{j+t+\lambda+k}\right) * x_{j+k+t+k+\lambda} =$$
$$= x_{j+t+\lambda} * x_{j+k+t+k+\lambda} =$$
$$= x_i * x_{i+2k}.$$

Then $\left[\left[x_1^{m+k}\right] x_{m+k+1}^{m+2k}\right]_i = \left[x_1^j \left[x_{j+1}^{j+m+k}\right] x_{j+m+k+1}^{m+2k}\right]_i$, for any $i \in \mathbf{N_m}$, $0 \le j \le k$. So $(Q;[\ ])$ is an $(m+k,m)-$semigroup, when $k < m$.

According to Proposition 1.2 $(Q;[\ ])$ is an $(m+k,m)-$rectangular band, when $k < m$.

It is clear that $B \subseteq Q$. Let $u \in Q$ and let $[B]$ be an $(m+k,m)-$subsemigroup of $(Q;[\ ])$ generated by B. Then, for $c \in B$ we have $u = ab = a*b = \left[\overset{i-1}{c}\ a\ \overset{k-1}{c}\ b\ \overset{m-i}{c}\right]_i \in [B]$, i.e. $Q \subseteq [B]$. So, $(Q;[\ ])$ is generated by B.

Let $(S;[\ ]')$ be an $(m+k,m)-$rectangular band when $k < m$ and let $f : B \to S$ be a map. By Proposition 1.2, there is a rectangular band $(S;\circ)$ such that $\left[x_1^{m+k}\right]_i' = x_i \circ x_{i+k}$, $x_1^{m+k} \in S^{m+k}$, $i \in \mathbf{N_m}$. Since $(Q;*)$ is a free rectangular band with a basis B, there is a homomorphism $g : Q \to S$, such that $g(b) = f(b), b \in B$.

Let $x_1^{m+k} \in Q^{m+k}$. Then

$$g\left(\left[x_1^{m+k}\right]_i\right) = g\left(x_i * x_{i+k}\right) = g(x_i) \circ g(x_{i+k}) = \left[g(x_1)g(x_2)\ldots g(x_{m+k})\right]_i'.$$

So, g as an extension of the map f, is $(m+k,m)-$homomorphism from $(Q;[\ ])$ into $(S;[\ ]')$. Hence $(Q;[\ ])$ is a free $(m+k,m)-$rectangular band with a basis B when $k < m$. ■

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FUNCTIONS PRESERVING PATH CONNECTEDNESS AND COMPACTNESS

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Abstract: *A function* $f:X\to Y$ *is preserving (path) connectedness if an image of an arbitrary (path) connected set is (path) connected. A function* $f:X\to Y$ *is preserving compactness if an image of an arbitrary compact set is compact. Under different conditions on spaces* X *and* Y *the relations of these type of maps and the notion of continuity are investigated.*

Several authors ([1], [2], [3]) consider the functions preserving compactness and conectedness. In this paper we prove some theorems about functions preserving compactness and path conectedness.

Definition. The function $f:X\to Y$ is preserving compactness if the image of every compact subset of X is compact in Y.

Definition. The function $f:X\to Y$ is preserving path connectedness if the image of every path connected subset of X is path connected in Y.

Theorem 1. Let X is a locally compact T_3 space, and Y is a locally path connected T_2 space. Let $f:X\to Y$ is a function preserving compactness and $f(X)=Y$, $f^{-1}(\{y\})$ is compact set for every $y\in Y$. If $f^{-1}(C)$ is a path connected set, for every $C\subseteq Y$, then f e closed map.

Proof. Let A is a closed subset of X and let $y\in Y\setminus f(A)$. Then $f^{-1}(\{y\})\cap A=\varnothing$. Since X is locally compact and regular, and $f^{-1}(\{y\})$ is compact, there exists an open set U in X such as ∂U is compact space, $f^{-1}(\{y\})\subseteq U$, $\partial U\cap f^{-1}(\{y\})=\varnothing$ and $U\cap A=\varnothing$.

Then $f(\partial U)$ is a compact set and $f(\partial U)\cap\{y\}=\varnothing$. Space Y is a locally path connected and there exists an open and path connected set R such that $R\cap f(\partial U)=\varnothing$ and $y\in R$. The set $f^{-1}(R)$ is path connected and

$f^{-1}(R)\cap\partial U=\varnothing$ therefore $f^{-1}(R)\subseteq U$. Finally $R\cap f(A)=\varnothing$ which means that $f(A)$ is a closed set.

With **N** we denote the set of natural numbers and with **R** the set of real numbers.

Theorem 2. Let $f:\mathbf{R}\to Y$ is a surjection. If $f^{-1}(\{y\})$ is a compact and connected set, for every $y\in Y$, $f(K)$ is a compact set, for every compact set $K\subseteq X$, and X is a locally compact T_3 space, and Y is a locally path connected T_2 space, then $f^{-1}(B)$ is path connected set, for every path connected set $B\subseteq Y$.

Proof. Let B be a path connected in Y. Then $f^{-1}(B)$ is connected set in **R** ([3] Theorem 1), so $f^{-1}(B)$ is an interval i.e.the set $f^{-1}(B)$ is path connected in **R**.

Theorem 3. If the Tychonoff space X is not locally connected at a point p, then there exist a function $f:X\to[0,1]$ preserving compactness and path connectedness witch is not continuous at p.

Proof. Let X is not locally connected at a point p. Then there is a neighborhood U of p such that: if K is a component of p in U, then K is not a neighborhood of p. Since X is Tychonoff space, there is a neighbourhood $V\subseteq U$ at p and continuous function $\overline{f}:X\to[0,1]$ such that $\overline{f}\,|_V=0$ and $\overline{f}\,|_{X\setminus U}=1$. Also exist a continuous function $g:X\to[0,1]$ such that $g(p)=1$ and $g\,|_{X\setminus V}=0$. Let $f(x)=\begin{cases}\overline{f}(x)+g(x), & x\in K\\ \overline{f}(x), & x\in X\setminus K\end{cases}$.

We will proof that f is a map from X to $[0,1]$. If $x\in V$, then $\overline{f}(x)=0$ so $f(x)=g(x)$. If $x\in X\setminus V$, then $g(x)=0$, so $f(x)=\overline{f}(x)$. In every case $0\le f(x)\le 1$.

Every neighbourhood at point p contains point y from $V\setminus K$ (because K is not a neighbourhood of p). Since $f(p)=1$ and $f(y)=\overline{f}(y)=0$ it follows that f is not a continuous at p.

We will proof that the image of compact subset of X is compact in $[0,1]$. Let $C\subseteq X$ is a compact set. Restriction of f on $F=(X\setminus V)\cup K$ (who is closed in X) is continuous ($g\,|_{X\setminus V}=0$ and on F holds $f=\overline{f}+g$). Now $C'=f(C\cap F)$ is compact set. But $f(x)=0$, for every $x\in V\setminus K$, so $f(C)=C'$ or $f(C)=C'\cup\{0\}$. Therefore $f(C)$ is compact set.

Let C is a path connected subset of X. Function f is contonous in $X\setminus K$ (because $f=\overline{f}$), so we may asume that $C\cap K\neq\varnothing$ (in contrary $f(C)$ is path connected set because of the continuity of f on $X\setminus K$). Similar, f is continuous on $X\setminus V$ (where

$f=\overline{f}$), and we assume that $C\cap V\neq\varnothing$. If $C\subseteq U$, then also $C\subseteq K$ (K is a component of connectedness on U, C is connected and $C\cap K\neq\varnothing$). But, f is continuous on K, and $f(C)$ is path connected.

It remains the case where $C\cap V\neq\varnothing$ and $C\cap(X\setminus U)\neq\varnothing$. In this case $\overline{f}(C)=[0,1]$ ($\overline{f}$ is continuous, $\overline{f}(V)=\{0\}$, $\overline{f}(X\setminus U)=\{1\}$ and C is path connected). Now, let, x is a point such that $\overline{f}(x)\neq 0$. It is clear that $x\notin V$. If $x\notin K$, then $f(x)=\overline{f}(x)$. If $x\in K\setminus V$, then $f(x)=\overline{f}(x)+g(x)=\overline{f}(x)+0=\overline{f}(x)$. This means that $f(x)=\overline{f}(x)$, for every $x\in X$ and $\overline{f}(x)\neq 0$. Because of that $f(C)\supseteq(0,1]$, and it is a path connected set.

Definition. Function $f:X\to Y$ is locally constant in $x\in X$ if there is a neighbourhood U on x such that f is constant on U.

Proposition 1. Let $f:X\to Y$ is a function such that image of arbitrary convergent sequence (x_n) in X is compact set in Y and X and Y are T_2 spaces. Then set $\{f(x_n)|n\in\mathbf{N}\}$ is finite, or $(f(x_n))$ is convergent sequence.

Proof. Let $\{f(x_n)|n\in\mathbf{N}\}$ be infinite and there exist at least two points of acumulation y_1 and y_2 (existing of this points refers to compactness of $f\left(\overline{\{x_n|n\in\mathbf{N}\}}\right)$) and the sequence (x_n) converges to x. Let $N=\{x_n|n\in\mathrm{N}\}\setminus f^{-1}(\{y_1\})$. Then $f(N)=\{f(x_n)|n\in\mathbf{N}\}\setminus\{y_1\}$. This means that $y_1\in\overline{\{f(x_n)|n\in\mathbf{N}\}}\setminus\{f(x_n)|n\in\mathbf{N}\}$. Similarly $y_1\in f\left(\overline{\{x_n|n\in\mathbf{N}\}}\right)$, so $y_1\in f\left(\overline{\{x_n|n\in\mathbf{N}\}}\right)\setminus\{f(x_n)|n\in\mathbf{N}\}$.

There exists $z\in\overline{\{x_n|n\in\mathbf{N}\}}\setminus\{x_n|n\in\mathbf{N}\}$ such that $f(z)=y_1$. Since X is T_2 space, $\overline{\{x_n|n\in\mathbf{N}\}}\setminus\{x_n|n\in\mathbf{N}\}=\{x\}$, so $f(x)=y_1$.

Similarly we obtain $f(x)=y_2$, and it follows $y_1=y_2$.

Proposition 2. Let $f:X\to Y$ is a function preserving path connectedness, X a locally path connected space, and Y a T_1.space. If f is not locally constant in $p\in X$, then $f(U)\cap V$ is infinite set, for every neighbourhood U on p and every neighbourhood V of $f(p)$.

Proof. Let U is a path connected neighbourhood on p. Then $f(U)$ is path connected and contains at least two points. Let V be an arbitrary neighbourhood on$f(p)$. Let as assume that the set $f(U)\cap V$ is finite. Because Y is T_1, we could choise a neighbourhood on $f(p)$ which do not contains other points of $f(U)$. According to this $f(p)$ is isolated point for $f(U)$. But $f(U)$ is path connected (and is connected also) and therefore the set does not contains isolated points.

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H^p, $p>1$ AS 2-NORMED SPACE

Vesna Manova-Erakovik

Abstract. *The notion of 2-norm is introduced in [2]. In [3] is given an equivalent definition by which the number of axioms is reduced to the same number of axioms as in the usual norm of a vector space. A class of functional is considered in [4], where given vector space is 2-normed and at the same time it is normed, with special relation between the 2-norm and the norm. In this paper we will give one 2-norm of the space* $H^p(U)$, $p>1$*, where* U *is the open unit disk.*

1. INTRODUCTION

We will consider a vector space X over a field Φ and $\dim X>1$. The field Φ is the field of the real numbers or the field of complex numbers.

Definition 1. Let X be a vector space over a field Φ and $\dim X>1$. The mapping $\|\cdot,\cdot\|: X^2\to\mathbf{R}$, which satisfies the following conditions

i) $\|x_1,x_2\|=0$ if and only if $\{x_1,x_2\}$ is linearly dependant set in X,

ii) $\|x_1,x_2\|=\|x_2,x_1\|$, $x_1,x_2\in X$,

iii) $\|\alpha x_1,x_2\|=|\alpha|\,\|x_1,x_2\|$, $\alpha\in\Phi$, $x_1,x_2\in X$,

iv) $\|x_1+x_1',x_2\|\le\|x_1,x_2\|+\|x_1',x_2\|$, $x_1,x_1',x_2\in X$.

is called 2-norm of the vector space X, and the ordered pair $(X,\|\cdot,\cdot\|)$ is called 2-normed space.

It is easy to prove that the 2-norm is nonnegative. As a consequence of the definition, is, also, the following property.

Lemma 1. Let X be 2-normed space. For any $x,y\in X$ and for each scalar $\alpha\in\Phi$ holds $\|x,y\|=\|x,y+\alpha x\|$.

Using the lemma 1, we get that for every matrix $A \in M_2(\Phi)$ and for every $(x_1,x_2) \in X^2$ holds $\| A\cdot(x_1,x_2)^T \| = |\det A| \, \| x_1,x_2 \|$, where $A\cdot(x_1,x_2)^T$ is the operation which is the same as the one when a matrix multiplies vector column.

Using the last equality, in [3], is given equivalent definition of a 2-norm.

Definition 2. Let X be a vector space over a field Φ with $\dim X > 1$. The mapping $\| \cdot,\cdot \| : X^2 \to \mathbf{R}$ which satisfies the following conditions:

$(P1)$ If $\| x,y \| = 0$ then the set of vectors $\{ x,y \}$ is linearly dependant,

$(P2)$ $\left\| A(x,y)^T \right\| = |\det A| \cdot \| x,y \|$, $x,y \in X$, $A \in M_2(\Phi)$.

$(P3)$ $\| x+x',y \| \le \| x,y \| + \| x',y \|$.

is called 2-norm, and the ordered pair $(X, \|\cdot,\cdot\|)$ is called 2-normed space.

Most of the results on functionals on 2-normed spaces are related to the class of bilinear functionals. One subclass, that is the class of alternative 2-forms, suits most to the properties of the 2-norm and is considered in [4].

The consideration of the alternative 2-forms is even more justified because of the motivation to introduce a class of bounded linear forms on 2-normed space. By the analogy to the definition of bounded linear functional, the bilinear functional is bounded if there exists positive constant K such that $|\Lambda(x,y)| \le K \| x,y \|$. For the linearly dependant $\{x,y\}$ holds $\| x,y \| = 0$, so we get $|\Lambda(x,y)| \le K \| x,y \| = 0$, i.e. $|\Lambda(x,y)| = 0$ i.e. $\Lambda(x,y) = 0$. This means that the class of bounded bilinear functionals with respect to the 2-norm consists of those functionals which are zero for all the pairs (x,y) such that $\{x,y\}$ is linearly dependant set. For this reason and because of the linearity, instead of bounded bilinear functionals, we will consider the class of alternative 2-linear forms. That is why the alternative 2- linear forms are called 2-linear functionals. We will consider the class of alternative linear 2-forms and use the following definition.

Definition 3. Let X be a vector space with $\dim X > 1$. The mapping $\Lambda : X \times X \to \Phi$ which satisfies the conditions

$$\Lambda(x+y,z) = \Lambda(x,z) + \Lambda(y,z),$$

$$\Lambda(A(x,y)^T) = (\det A)\Lambda(x,y),$$

Is called 2-linear functional.

It is not hard to prove that this definition is equivalent to the definition of the alternative 2-linear form.

2. SOME PROPERTIES OF $L^p(T)$

Let T be the boundary of the unit disk. The properties of the normed space $L^p(T), p > 1$ as 2-normed space are of great importance for 2-norming the normed vector space $H^p(U) = H^p$, $p > 1$. In order to have complete and clear understanding of the subsequent part, and because of the great importance of H^p as 2-normed space, we will, partially, give the way of 2-norming $L^p(T), p > 1$.

In this part we will consider one property of the space $L^p(T)$ that will help us to introduce 2-norm on $H^p, p > 1$. Let

$$\| f^*, g^* \| = \left\{ \int\limits_{T\times T} \left| \begin{vmatrix} f^*(\theta) & f^*(\theta_1) \\ g^*(\theta) & g^*(\theta_1) \end{vmatrix} \right|^p d\theta d\theta_1 \right\}^{1/p} \text{, for } f^*, g^* \in L^p(T).$$

Using

(1) $$\left| \begin{vmatrix} f^*(\theta) & f^*(\theta_1) \\ g^*(\theta) & g^*(\theta_1) \end{vmatrix} \right|^p \leq 2^{\,p-1}\left(| f^*(\theta) g^*(\theta_1) |^p + | g^*(\theta) f^*(\theta_1) |^p \right)$$

it is not hard to prove that $\|\cdot,\cdot\|\, L^p(T) \times L^p(T) \to \text{Ў}^{+} \cup \{0\}$ is well defined function and that $\| f^*, g^* \| \leq 2 \| f^* \| \| g^* \|$, where $\| f^* \|$ and $\| g^* \|$ stand for the usual norm of f^*, g^*, respectively, in $L^p(T)$.

Lemma. Let $f^*, g^* \in L^p(T)$ and $\| f^*, g^* \| = 0$. Then there exists scalar α such that $f^*(\theta) = \alpha g^*(\theta)$ a.e. on T or $g^*(\theta) = \alpha f^*(\theta)$ a.e. on T.

Proof. Let us assume that $\| f^*, g^* \| = 0$ and $f^* \neq 0$ and $g^* \neq 0$. Then $\begin{vmatrix} f^*(\theta) & f^*(\theta_1) \\ g^*(\theta) & g^*(\theta_1) \end{vmatrix} = 0$ a.e. with respect to the direct product of the Lebesque measure on $T \times T$. The sets $U_{f^*} = \{\theta / \theta \in [0, 2\pi) / f^*(\theta) \neq 0\}$ and $U_{g^*} = \{\theta / \theta \in [0, 2\pi) / g^*(\theta) \neq 0\}$ have positive measure, i.e. $m(U_{f^*}) > 0$ and $m(U_{g^*}) > 0$. On the other hand $W_{f^*,g^*} = \{(\theta, \theta_1) / f^*(\theta) g^*(\theta_1) - g^*(\theta) f^*(\theta_1) \neq 0\}$ is a set with measure zero, i.e. $(m \times m)(W_{f^*,g^*}) = 0$, where m is the Lebesque measure on T.

If we assume that $m(U_{f^*} \cap U_{g^*}) = 0$, then $U_{f^*} \cap U_{g^*} = \varnothing$ a.e. with the respect to the Lebesque measure. According to this, for arbitrary $(\theta, \theta_1) \in U_{f^*} \times U_{g^*}$ holds $f^*(\theta) \neq 0, g^*(\theta_1) \neq 0$ and $g^*(\theta) = 0, f^*(\theta_1) = 0$. Then $f^*(\theta) g^*(\theta_1) - g^*(\theta) f^*(\theta_1) = f^*(\theta) g^*(\theta_1) \neq 0$. Thus $(\theta, \theta_1) \in W_{f^*,g^*}$, i.e. $U_{f^*} \times U_{g^*} \subseteq W_{f^*,g^*}$. Using the properties of the measure $m \times m$ we get that $(m \times m)(W_{f^*,g^*}) \geq (m \times m)(U_{f^*} \times U_{g^*}) = m(U_{f^*}) m(U_{g^*}) > 0$, which is in contradiction with the assumption $(m \times m)(W_{f^*,g^*}) = 0$. Therefore

$$m(U_{f^*} \cap U_{g^*}) > 0.$$

We will consider the case $m(U_{f^*} \cap U_{g^*}) > 0$ and $m(U_{f^*} \setminus U_{g^*}) > 0$. The case $m(U_{f^*} \cap U_{g^*}) > 0$ and $m(U_{g^*} \setminus U_{f^*}) > 0$ is analogous.

If $(\theta, \theta_1) \in (U_{f^*} \setminus U_{g^*}) \times (U_{f^*} \cap U_{g^*})$, then $\theta \in U_{f^*}, \theta \notin U_{g^*}, \theta_1 \in U_{f^*}, \theta_1 \notin U_{g^*}$ and $f^*(\theta) g^*(\theta_1) - g^*(\theta) f^*(\theta_1) = f^*(\theta) g^*(\theta_1) \neq 0$.

So $(U_{f*} \setminus U_{g*}) \times (U_{f*} \cap U_{g*}) \subseteq W_{f*,g*}$

$(m\times m)(W_{f*,g*}) \geq (m\times m)[(U_{f*} \setminus U_{g*}) \times (U_{f*} \cap U_{g*})] = m(U_{f*} \setminus U_{g*}) m(U_{f*} \cap U_{g*}) > 0$ which contradicts the assumption. Therefore $m(U_{f*} \setminus U_{g*}) = 0$, and similarly

$$m(U_{g*} \setminus U_{f*}) = 0.$$

From all the above, we get that $U_{f*} = U_{g*} = V$ a.e. with respect to the Lebesque measure on T, so $m(V) > 0$.

For $\theta \in U$, let $U_\theta = \{\theta_1 / \theta_1 \in V, f*(\theta) g*(\theta_1) - g*(\theta) f*(\theta_1) \neq 0\}$.Let us assume that for each $\theta \in V$, $m(U_\theta) > 0$. Since $f* = g* = 0$ a.e. on $T \setminus V$ with the respect to the Lebesque measure, we get that $f* = g* = 0$ a.e. on $T \setminus U_\theta$ with the respect to the Lebesque measure and

$$h(\theta) = | f*(\theta) g*(\theta_1) - g*(\theta) f*(\theta_1) |^p > 0 \text{ on } U_\theta .$$

Then

$$\varphi(\theta) = \int_T | f*(\theta) g*(\theta_1) - g*(\theta) f*(\theta_1) |^p d\theta_1 =$$

$$= \int_{U_\theta} | f*(\theta) g*(\theta_1) - g*(\theta) f*(\theta_1) |^p d\theta_1 > 0$$

and therefore $\int_{T\times T} \left| \begin{vmatrix} f*(\theta) & f*(\theta_1) \\ g*(\theta) & g*(\theta_1) \end{vmatrix} \right|^p d\theta d\theta_1 = \int_T \varphi(\theta) d\theta = \int_U \varphi(\theta) d\theta > 0$.

The last inequality is in contradiction with the assumption that $\| f*, g* \| = 0$. Therefore, there exists $\theta_o \notin V$ such that $f*(\theta_o) g*(\theta_1) - g*(\theta_o) f*(\theta_1) = 0$ a.e. V. Since $f*(\theta_o), g*(\theta_o) \neq 0$, we have

(2) $$g*(\theta_1) = \frac{g*(\theta_o)}{f*(\theta_o)} f*(\theta_1), \text{ for each } \theta_1 \in V .$$

On the other hand, $f*, g* = 0$ on $T \setminus V$, so the equality (2) holds on T. Therefore, $g* = \alpha f*$ on T, where $\alpha = \dfrac{g*(\theta_o)}{f*(\theta_o)}$.

Using the introduced function, which is 2-norm on $L^p(T), p > 1$, (see $[5]$) in the subsequent part we will define 2-norm on $H^p(U), p > 1$.

3. $H^p, p > 1$ AS 2-NORMED SPACE

Using the introduced function on $L^p(T) \times L^p(T)$ we will define mapping on the set $H^p(T) \times H^p(T), p > 1$ which is 2-norm on H^p.

It is known that, for any function $f \in H^p$ the boundary value $f^*(\theta) = \lim_{r \to 1} f(re^{i\theta})$ exists a.e. on $[0, 2\pi]$ and belongs to $L^p(T)$. (see $[1]$) For $f, g \in H^p$ let

(3) $$\| f, g \| = \| f^*, g^* \|$$

Because of the uniqueness of f^* and g^*, the mapping (3) is well defined. We will show that it is also a 2-norm on $H^p(U), p > 1$.

$i)$ If $f, g \in H^p$ are such that $f = \alpha g$ for some scalar α then, using the properties of the boundary values we get that $f^* = \alpha g^*$. So,

$$\| f, g \| = \| f^*, g^* \| = \| \alpha g^*, g^* \| = | \alpha | \| g^*, g^* \| = | \alpha | \cdot 0 = 0.$$

Conversely, let $\| f, g \| = 0$. Then because of (3) $\| f^*, g^* \| = 0$, so there exists scalar $\alpha \in \mathrm{J}$ such that $f^* = \alpha g^*$ a.e. Hence,

$f^*(e^{i\theta}) P(r, t - \theta) = \alpha g^*(e^{i\theta}) P(r, t - \theta)$ a.e., where $P(r, \theta)$ is the Poisson kernel (see $[1]$). Therefore

$$\frac{1}{2\pi} \int_0^{2\pi} f^*(e^{i\theta}) P(r, t - \theta) d\theta = \alpha \frac{1}{2\pi} \int_0^{2\pi} g^*(e^{i\theta}) P(r, t - \theta) d\theta, \text{ i.e. } f(z) = \alpha g(z).$$

$ii)$ Let $A \in M_2(\mathbf{C})$ be arbitrary chosen and let $f, g \in H^p(U), p > 1$. Then, using the properties of determinants, we get

$$\| A(f, g)^T \| = \| A(f^*, g^*)^T \| = \left\{ \int_{T \times T} | A(f^*, g^*)^T |^p \, d\theta d\theta_1 \right\}^{1/p} =$$

$$= \left\{ \int_{T \times T} \left| \begin{vmatrix} a_{11} f^*(\theta) + a_{12} g^*(\theta) & a_{11} f^*(\theta_1) + a_{12} g^*(\theta_1) \\ a_{21} f^*(\theta) + a_{22} g^*(\theta) & a_{21} f^*(\theta_1) + a_{22} g^*(\theta_1) \end{vmatrix} \right|^p d\theta d\theta_1 \right\}^{1/p} =$$

$$= \left\{ \int_{T \times T} \left| \begin{vmatrix} f^*(\theta) & f^*(\theta_1) \\ g^*(\theta) & g^*(\theta_1) \end{vmatrix} \begin{vmatrix} a_{11} & a_{12} \\ a_{21} & a_{22} \end{vmatrix} \right|^p d\theta d\theta_1 \right\}^{1/p} = | \det A | \| f, g \|$$

$iii)$ Let $f, g, h \in H^p(U)$ are arbitrary chosen. Using the inequality (1) we get

$$\|f+g,h\|=\|f^*+g^*,h^*\|=\left\{\int_{T\times T}\left|\begin{vmatrix} f^*(\theta)+g^*(\theta) & f^*(\theta_1)+g^*(\theta_1) \\ h^*(\theta) & h^*(\theta_1)\end{vmatrix}\right|^p d\theta d\theta_1\right\}^{1/p}=$$

$$\leq\left\{\int_{T\times T}\left(\left|\begin{vmatrix} f^*(\theta) & f^*(\theta_1) \\ h^*(\theta) & h^*(\theta_1)\end{vmatrix}\right|+\left|\begin{vmatrix} g^*(\theta) & g^*(\theta_1) \\ h^*(\theta) & h^*(\theta_1)\end{vmatrix}\right|\right)^p d\theta d\theta_1\right\}^{1/p}\leq$$

$$\leq\left\{\int_{T\times T}\left|\begin{vmatrix} f^*(\theta) & f^*(\theta_1) \\ h^*(\theta) & h^*(\theta_1)\end{vmatrix}\right|^p d\theta d\theta_1\right\}^{1/p}+\left\{\int_{T\times T}\left|\begin{vmatrix} g^*(\theta) & g^*(\theta_1) \\ h^*(\theta) & h^*(\theta_1)\end{vmatrix}\right|^p d\theta d\theta_1\right\}^{1/p}=$$

$$=\|f^*,h^*\|+\|g^*,h^*\|=\|f,h\|+\|g,h\|$$

So we proved that $(H^p,\|\cdot,\cdot\|)$ is 2-normed space.

The expected result is that it holds the following equality

$$\sup_{0\leq r,r_1<1}\left\{\int_0^{2\pi}\left|\begin{vmatrix} f(re^{i\theta}) & f(r_1e^{i\theta_1}) \\ g(re^{i\theta}) & f(r_1e^{i\theta_1})\end{vmatrix}\right|^p d\theta d\theta_1\right\}^{\frac{1}{p}}=\|f,g\|,$$

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TOPOLOGICAL PROPERTIES OF THE PARETO-OPTIMAL SET IN VECTOR OPTIMIZATION PROBLEM

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Abstract: *In this paper we consider the Pareto-optimal set in vector optimization problem with concave constraint and objective functions. It is shown the problem of construction of a retraction from the feasible domain onto the Pareto-optimal set, if one of objective functions is strictly concave. Using this theoretical result it is proved that the Pareto-optimal set is compact and contractible, and has the fixed point properties.*

Keywords: *Pareto-optimal, topological structure, convex, compact, contractible, fixed point property.*

1. INTRODUCTION

The standard form of the vector maximization problem is to find a variable $x(x_1, x_2,..., x_m) \in R^m$, $m \geq 1$, so as to

maximize $F(x) = (f_1(x), f_2(x),..., f_n(x))$

subject to $G(x) = (g_1(x), g_2(x),..., g_p(x)) \geq 0$

$x_i \in [a_i, b_i] \forall i \in J_m$,

where $\{f_i\}_{i=1}^n$ are given objective functions, $n \geq 2$, $\{g_i\}_{i=1}^p$ are given inequality constraint functions, $p \geq 1$, and a_i and b_i are the lower and upper bounds for x_i, $a_i < b_i$, and $J_m = \{1,2,..., m\}$ is the index set.

Let the feasible domain denote

$$X = \{x(x_1, x_2,..., x_m) \in R^m \mid g_i(x) \geq 0 \forall i \in J_p, a_i \leq x_i \leq b_i \forall i \in J_m\}.$$

As usual, let us assume that the set X is nonempty.

Definition 1. A point $x \in X$ is called Pareto-optimal solution if and only if there does not exist a point $y \in X$ such that $f_i(y) \geq f_i(x)$ for all $i \in J_n$ and $f_k(y) > f_k(x)$ for some $k \in J_n$. The set of the Pareto-optimal solutions of X is denoted by $Max(X, F)$ and is called Pareto-optimal set.

Considering topological properties of the Pareto-optimal set is started by [8], see also [3], [6] and [9]. The well-known open problems in optimization are the compactness, contractibility and fixed point properties of the Pareto-optimal set. Compactness of this set is studied in [6] and [11]. Contractibility of the Pareto-optimal set is considered in [1], [5] and [6]. Fixed point properties of the Pareto-optimal set are discussed in [10].

In this paper, let the given functions $\{f_i\}_{i=1}^n$ and $\{g_i\}_{i=1}^p$ all be continuous and concave. As a result we find that the feasible domain X is convex. Let us also assume that a function f_λ of $\{f_i\}_{i=1}^n$ is strictly concave.

2. GENERAL DEFINITIONS AND NOTIONS

Let a function $dis: X \times X \to R_+$ be a metric. In (X, dis), let τ be the topology induced by dis. In (X, τ), for $Y \subset X$ we recall:

Definition 2. The set Y is a retract of X if and only if there exists a continuous function $r: X \to Y$ such that $r(x) = x$ for all $x \in Y$. The function r is called a retraction.

Definition 3. A continuous function $d: X \times [0;1] \to X$ is a deformation retraction of X onto Y if and only if $d(x,0) = x$, $d(x,1) \in Y$, and $d(a,t) = a$ for all $x \in X$, $a \in Y$, and $t \in [0;1]$. The set Y is called a deformation retract of X.

Definition 4. The set Y is contractible if and only if there exist a continuous function $c: Y \times [0;1] \to Y$ and $a \in Y$ such that $c(x,0) = a$ and $c(x,1) = x$ for all $x \in Y$.

Compactness and contractibility of sets are preserved under retraction. This means that the following statements are true: If the set X is compact and Y is a retract of X, then set Y is compact too; If the set X is contractible and Y is a retract of X, then set Y is contractible too.

It is known that: convexity implies contractibility, contractibility implies path-connectedness and path-connectedness implies connectedness. But, in general the converse does not hold [4] [6].

Definition 5. The topological space X is said to have a fixed point property if and only if every continuous function $h: X \to X$ from this set into itself has a fixed point, i.e. there is a point $x \in X$ such that $x = h(x)$.

Let us consider a point-to-set mapping $\varphi: X \Rightarrow X$. Let it be upper semi-continuous with nonempty, compact and convex images, shortly we say that φ is *cusco*.

Definition 6. The topological space X is said to have a Katutani fixed point property if and only if every *cusco* $\varphi: X \Rightarrow X$ has a fixed point, i.e. there is a point $x \in X$ such that $x \in \varphi(x)$.

Fixed point and Kakutani fixed point properties of sets are preserved under retraction. This means that the following statements are true: If the set X has the fixed point property and Y is a retract of X, then set Y has the fixed point property too; If the set X has the Kakutani fixed point property and Y is a retract of X, then set Y has the Kakutani fixed point property too.

We introduce the following notations: for $x, y \in R^n$: $x(x_1, x_2, ..., x_n) \underline{\phi}\, y(y_1, y_2, ..., y_n)$ means $x_i \geq y_i$ for all $i \in J_n$ and $x_k > y_k$ for some $k \in J_n$.

3. CONSTRUCTING THE RETRACTION

Define a function $f: X \to R$ by $f(x) = \sum_{j=1}^{n} f_j(x)$ for all $x \in X$.

Define also a point-to-set mapping $\psi: X \Rightarrow X$ by $\psi(x) = \{y \in X \mid F(y) \geq F(x)\}$ for all $x \in X$. It can be shown that $\psi(x)$ is a nonempty, compact and convex set for all $x \in X$.

These notes allow presenting the main theorem of this section.

Theorem 1. There exists a retraction $r: X \to Max(X,F)$ such that $r(X) = Max(X,F)$ and $r(x) = Arg\max(f, \psi(x))$ for all $x \in X$.

Lemma 1. *If* $x \in X$, *then* $|Arg\max(f, \psi(x))| = 1$ *and* $Arg\max(f, \psi(x)) \subset Max(X,F)$.

Proof. As described earlier, the function f is strictly concave and $\psi(x)$ is convex, therefore we obtain $\left|Arg\max(f,\psi(x))\right| = 1$.

Let choose $y \in Arg\max(f,\psi(x))$ and assume $y \notin Max(X,F)$. From $y \notin Max(X,F)$ it follows that there exists $z \in X$ such that $F(z)\underline{\phi}F(y)$. As a result we derive $z \in \psi(x)$ and $f(z) > f(y)$. This leads to a contradiction, therefore we obtain $y \in Max(X,F)$. The lemma is proved.

Using the results of Lemma 1 we are in a position to construct a function $r : X \to Max(X,F)$ such that $r(x) = Arg\max(f,\psi(x))$ for all $x \in X$.

Lemma 2. If $x \in X$, $x \in Max(X,F)$ is equivalent to $\{x\} = \psi(x)$.

Proof. Let $x \in Max(X,F)$ and assume that $\{x\} \neq \psi(x)$. From $x \in \psi(x)$ and $\{x\} \neq \psi(x)$, it follows that there exists $y \in \psi(x) \setminus \{x\}$ such that $F(y) \geq F(x)$. Let us choose $t \in (0;1)$ and $z = tx + (1-t)y$; then $z \in \psi(x)$. But $x \neq y$ implies $f_\lambda(z) > f_\lambda(x)$, which contradicts $x \in Max(X,F)$. Then we obtain $\{x\} = \psi(x)$.

Conversely, let $\{x\} = \psi(x)$ and assume $x \notin Max(X,F)$. From the assumption $x \notin Max(X,F)$, it follows that there exists $y \in X$ such that $F(y)\underline{\phi}F(x)$. Thus we deduce that $y \in \psi(x)$ and $x \neq y$, which contradicts the condition $\{x\} = \psi(x)$. Then we obtain $x \in Max(X,F)$. The lemma is proved.

Lemma 3. $r(X) = Max(X,F)$.

Proof. From Lemma 1 it follows that $r(X) \subset Max(X,F)$. Applying Lemma 2 we deduce $r(Max(X,F)) = Max(X,F)$. This means that $r(X) = Max(X,F)$. The lemma is proved.

Lemma 4. The point-to-set mapping ψ is continuous on X.

Proof. First, we will prove that if $\{x_k\}_{k=1}^{\infty}, \{y_k\}_{k=1}^{\infty} \subset X$ is a pair of sequences such that $\lim_{k\to\infty} x_k = x_0 \in X$ and $y_k \in \psi(x_k)$ for all $k \in N$, then there exists a convergent subsequence of $\{y_k\}_{k=1}^{\infty}$ whose limit belongs to $\psi(x_0)$.

The assumption $y_k \in \psi(x_k)$ for all $k \in N$ implies $F(y_k) \geq F(x_k)$ for all $k \in N$. From $\{y_k\}_{k=1}^{\infty} \subset X$ it follows that there exists a convergent subsequence $\{q_k\}_{k=1}^{\infty} \subset \{y_k\}_{k=1}^{\infty}$ such that $\lim_{k\to\infty} q_k = y_0 \in X$. Therefore, there exists a convergent subsequence $\{p_k\}_{k=1}^{\infty} \subset \{x_k\}_{k=1}^{\infty}$ such that $q_k \in \psi(p_k)$ and $\lim_{k\to\infty} p_k = x_0$. Thus, we find that $F(q_k) \geq F(p_k)$ for all $k \in N$. Taking the limit as $k \to \infty$ we obtain $F(y_0) \geq F(x_0)$. This implies $y_0 \in \psi(x_0)$. This means that ψ is upper semi-continuous on X [7].

Second, we will prove that if $\{x_k\}_{k=1}^{\infty} \subset X$ is a convergent sequence to $x_0 \in X$ and $y_0 \in \psi(x_0)$, then there exists a sequence $\{y_k\}_{k=1}^{\infty} \subset X$ such that $y_k \in \psi(x_k)$ for all $k \in N$ and $\lim_{k\to\infty} y_k = y_0$.

As usual, let us denote the distance between the point $y_0 \in X$ and the set $\psi(x_k) \subset X$ by $d_k = \inf\{dis(y_0, x) \mid x \in \psi(x_k)\}$. There are two cases as follows: if $y_0 \in \psi(x_k)$, then let $y_k = y_0$; if $y_0 \notin \psi(x_k)$, then let $y_k = \bar{y}$.

So we get a sequence $\{d_k\}_{k=1}^{\infty} \subset R_+$ and a sequence $\{y_k\}_{k=1}^{\infty} \subset X$ such that $y_k \in \psi(x_k)$ for all $k \in N$ and $d_k = dis(y_0, y_k)$. Since $\lim_{k\to\infty} x_k = x_0$, the sequence $\{d_k\}_{k=1}^{\infty}$ is convergent and $\lim_{k\to\infty} d_k = 0$. Finally, we obtain $\lim_{k\to\infty} y_k = y_0$. This means that the point-to-set mapping ψ is lower semi-continuous on X [7].

In summary, ψ is continuous on X. The lemma is proved.

Lemma 5 [12, Theorem 9.14]. Let $S \subset R^n$, $\Theta \subset R^m$, $g : S \times \Theta \to R$ a continuous function, and $D : \Theta \rightrightarrows S$ be a compact-valued and continuous point-to-set mapping. Then, the function $g^* : \Theta \to R$ defined by $g^*(\theta) = \max\{g(x,\theta) \mid x \in D(\theta)\}$ is continuous on Θ, and the point-to-set mapping $D^* : \Theta \rightrightarrows S$ defined by $D^*(\theta) = \{x \in D(\theta) \mid g(x,\theta) = g^*(\theta)\}$ is compact-valued and upper semi-continuous on Θ.

Lemma 6. The function r is continuous on X.

Proof. Applying Lemma 5 we derive that f is continuous on X. The point-to-set mapping ψ is compact-valued and continuous. According to Lemma 1 we deduce that r is upper semi-continuous point-to-point mapping. An upper semi-continuous point-to-point mapping is continuous when viewed as a function. In result, the function r is continuous on X. The lemma is proved.

We are now in a position to prove the main result of this section.

Proof of Theorem 1. From Lemmas 1, 3 and 6, it follows that there exists a continuous function $r : X \to Max(X, F)$ such that $r(X) = Max(X, F)$ and $r(x) = Arg\max(f, \psi(x))$ for all $x \in X$. The theorem is proved.

4. TOPOLOGICAL STRUCTURE OF THE PARETO-OPTIMAL SET

In this section, we will discuss the topological structure of the Pareto-optimal set using the theoretical result obtained above.

Theorem 2. The Pareto-optimal set $Max(X, F)$ is compact and contractible, and has the fixed point and the Kakutani fixed point properties.

The most well-known useful fixed point theorems are the following:

Lemma 7 [12, Theorem 9.31 - Schauder's Fixed Point Theorem]. Let $h : S \to S$ be continuous function from nonempty, compact and convex set $S \subset R^n$ into itself, then h has a fixed point.

Lemma 8 [12, Theorem 9.31 - Kakutani's Fixed Point Theorem]. Let $S \subset R^n$ be nonempty, compact and convex set and the point-to-set mapping $\varphi : S \rightrightarrows S$ be *cusco*, then φ has a fixed point.

Now, we are in a position to prove the main result of this section.

Proof of Theorem 2. We have shown that X is compact and contractible, and has the fixed point and the Kakutani fixed point properties, see Lemmas 7 and 8. From Theorem 1 it follows that $Max(X, F)$ is a retract of X. As a result we obtain that

$Max(X, F)$ is compact and contractible, and has the fixed point and the Kakutani fixed point properties. The theorem is proved.

Remark 1. The Kakutani fixed point property is very closely related to the fixed point property. If $S \subset R^n$ has the Kakutani fixed point property, then since any continuous point-to-point mapping can be viewed as a *cusco* it follows that the set S will also have the fixed point property.

Let $\varphi : S \Rightarrow S$ *be a point-to-set mapping and denote* $gph(\varphi) = \{(x, y) \in S \times S \mid y \in \varphi(x)\}$. *It is called the graph of* φ.

Remark 2. Let $S \subset R^n$ be compact. It can be shown that the set S having the Kakutani fixed point property is equivalent to S having the fixed point property. In Remark 1, we have shown that if S has the Kakutani fixed point property, then S has the fixed point property. Now, let S have the fixed point property and let φ be *cusco*. From Cellina's Theorem it follows that there is an approximate continuous selection h of φ [2, Theorem 8.2.5]. That is, for each $k \in N$ there exists a continuous function $h_k : S \to S$ such that $dis((x, h_k(x)), gph(\varphi)) < \frac{1}{k}$ for all $x \in S$. From S has the fixed point property it follows that h_k has a fixed point $x_k \in S$. As a result we get a sequence $\{x_k\}_{k=1}^{\infty} \subset S$ such that $dis((x_k, x_k), gph(\varphi)) < \frac{1}{k}$. The set S is compact implies that there exists a convergent subsequence $\{x'_{m(k)}\}_{m(k)=1}^{\infty} \subset \{x_k\}_{k=1}^{\infty}$ such that $\lim_{k\to\infty} x'_{m(k)} = x_0 \in S$. We also see that $dis((x'_{m(k)}, x'_{m(k)}), gph(\varphi)) < \frac{1}{m(k)}$. But if φ is *cusco*, then $gph(\varphi)$ is closed. Taking the limit as $k \to \infty$ we have $m(k) \to \infty$ and obtain $\lim_{k\to\infty}(x'_{m(k)}, x'_{m(k)}) = (x_0, x_0) \in gph(\varphi)$. This means that $x_0 \in S$ is the fixed point for φ.

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QUANTITATIVE STRUCTURE-SCAVENGING ACTIVITY RELATIONSHIP OF PHENOLIC COMPOUNDS

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Abstract: *A lot of investigations addressing correlation between antioxidant activity and theoretically generated descriptors can be found in literature. This task is quite ambitious, bearing in mind that the rather complicated interactions in the cell allow for different possible mechanisms of the radical scavenging reaction. In this investigation we tried to simplify the problem by looking for direct correlations between calculated characteristics and scavenging activity, neglecting the specificity of cellular environment. A set of 15 phenolic compounds and their phenoxyl radicals were investigated with the DFT method using UB3LYP/6-31+G(d,p). Some of obtained indices were related to the results of the DPPH scavenging activity.*

Keywords: *DFT calculations, DPPH scavenging activity, QSAR*

1. INTRODUCTION

The antioxidants are widely used as drugs and nutrition supplements. Their function consists in scavenging of active radicals generated in different ways in higher organisms. Thus they prevent undesired chemical changes in cells and the development of diseases like cancer, atherosclerosis, different inflammations etc. [8,16].

The scavenging reaction of phenolic antioxidants can be illustrated by the next scheme:

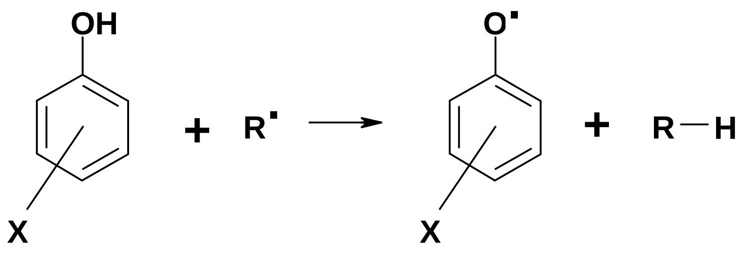


Fig. 1. Reaction between phenolic antioxidats and active radicals.

The parent antioxidant molecule releases a radical which has no potential for participation in unwanted chain radical processes but may take part in trapping of active radicals [2] named “second radical scavenging” (Fig. 2).

The reaction mechanism on the first step is arguable, since various alternatives are discussed in literature. Possible pathways are: direct H-atom transfer, single-electron transfer, sequential proton loss electron transfer, etc. [5, 11-13]

The antioxidant efficiency is determined by the rate and the degree of these reactions.

The experimentally determinable and the theoretically computable descriptors of antioxidant activity can be divided into the following groups: i) indices estimating O-H bond strength – bond dissociation enthalpy (BDE) [10] or structural parameters – bond distance,

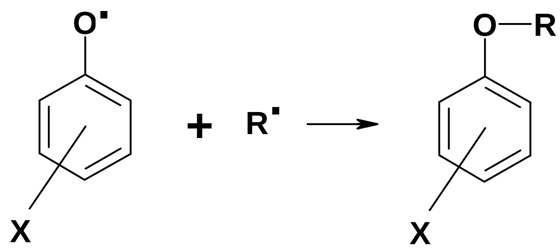


Fig. 2. Second radical scavenging reaction.

charge distributions etc.; ii) indices presenting molecular electron-donation capacity – ionization potential or its theoretical analog – HOMO energy; iii) indices showing the stability of obtained phenoxyl radicals (Ar-O•) – spin distribution, C-O bond distance in the radicals etc. [4].

Correlations between antioxidant activity and theoretically or experimentally derived descriptors belonging to the types mentioned above are usually obscured [19] by the dependence of the antioxidant activity on additional factors as permeability into the cell (lipophilicity), coordination ability and resistance to enzymatic degradation [1]. Therefore, a preceding analysis of correlation between the main descriptors and scavenging activity is more meaningful.

On the other hand, it is found that theoretical and experimental methods of the BDE determination give reliable results for monophenols, but not for diphenols [18]. The presence of hydrogen bond with the oxygen atom from the dissociable hydroxyl group facilitates considerably the dissociation process. This effect can be evaluated well if the experiment for antioxidant/scavenging activity determination is carried out in aprotic solvent. In this case the calculated BDE accounts reliably for the contribution of the adjacent hydroxyl group. In contrast, if the experiment of antioxidant/scavenging activity determination is performed in protic solvents, forming hydrogen bonds with the molecule, this effect can not be estimated purely by the theoretically obtained BDE for monophenols.

The establishment of relations between basic descriptors and antioxidant/scavenging activity is also hindered due to the different reaction mechanisms which take place.

In polar solvents more probable reaction mechanism is two-step electron-proton transfer, rather than one-step hydrogen atom transfer. In such a case not BDE of O-H bond but ionization potential (or HOMO energy) is the proper descriptor [5, 11-13].

The aim of the present investigation is to create quantum-chemical models of a series of natural antioxidants and their phenoxyl radicals. Secondly, comparison of their ability to react with DPPH according to the calculated structural indices of antioxidant/scavenging activity is attempted. Data from DPPH-tests found in the literature were used for the purpose [14]. The obtained results will be used for search of correlation between scavenging activity and structural indices.

2. METHODS

To carry out this study, we have selected a set of 15 phenolic natural antioxidants. The DPPH-scavenging activity has been taken from the Ref. [14] and is expressed as the percent of relative scavenging activity (%RSA) for 10-min reaction periods and 0.25 relative concentration of antioxidant compared to DPPH in mol/mol.

Full geometry optimization was performed with the unrestricted B3LYP method [15] using GAUSSIAN'03 program package [7] and 6-31+G(d,p) orbital basis set [6]. It was found that DFT methods and especially unrestricted B3LYP gives more reliable results for the BDE of O-H [3]. The utilization of other post HF methods is rather expensive. Usually the extension of the orbital basis does not improve the results [17].

All available intra-molecular hydrogen bond have been taken into account in the initial geometry generation. In case of two hydroxyl groups only *para*-hydroxyl group dissociation has been considered in the formation of corresponding radicals.

The linear regression was performed by Microsoft Excel program package.

3. RESULTS AND DISCUSSIONS

The values of the following most frequently used parameters were obtained: i) energies of highest occupied molecular orbital (E_{HOMO}) of compounds; ii) C-O bond distance (the O atom is from the dissociated hydroxyl group) in the respective radicals; iii) Mulliken atomic spin densities at the same oxygen atoms.

The descriptors mentioned above can be considered as random variables. We have investigated the hypothesis of linear dependence between the scavenging activities of the phenolic compounds (considered as dependent variables), obtained by DPPH tests [14] and each of the three parameters (i, ii, iii) considered as independent variables separately. The values of parameters to be analyzed are presented in Tab. 1.

Tab. 1. Values of %RSA and considered descriptors.

Phenolic compound	Percent Relative Activity (Y)	E_{HOMO} *(X)	C-O distance** (Z)	Mulliken atomic spin density at O atom (W)
Dihydrocaffeic acid	83.7	-0.22212	1.26005	0.346100
Rozmarinic acid	82.4	-0.22119	1.25225	0.285608
Caffeic acid	63.6	-0.22905	1.28427	0.380135
Chlorogenic acid	49.2	-0.22702	1.25282	0.286886
Sinapic acid	54.4	-0.22383	1.24528	0.315095
Ferulic acid	26.7	-0.22262	1.25145	0.315920
p-Coumaric acid	3.6	-0.23360	1.24970	0.335027
Hydroxytyrosol	56.5	-0.21946	1.26029	0.344569
Oleuropein	41.3	-0.22182	1.25993	0.340505
Tyrosol	2.6	-0.22555	1.25937	0.404905
α-Tocopherol	52.8	-0.19265	1.26025	0.351094
Trolox	53.4	-0.19752	1.25971	0.355192
TBHQ	52.3	-0.20985	1.25884	0.362101
BHA	16.1	-0.20644	1.25980	0.379353
BHT	5.7	-0.22943	1.25745	0.347618

- in eV; ** in Å.

In the second column of the Tab.1 %RSA is presented. In the following columns the respective values of the parameters (i)-(iii) are given.

The linear regression equation of the variable Y on the variables X, Z and W respectively are:

(1) $Y = 152.07 + 498.7X,\ r_{YX} = 0.22,\ s = 26.82,$

(2) $Y = -731.49 + 615.57Z,\ r_{YZ} = 0.20,\ s = 26.95,$

(3) $Y = 145.32 - 298.15W,\ r_{YW} = 0.37,\ s = 25.56.$

The low values of the correlation coefficients *r* in all three equations allow us to reject the hypotheses of linear correlation between the scavenging activity and the calculated parameters under study. The details of the correlation analysis are given in the Appendix.

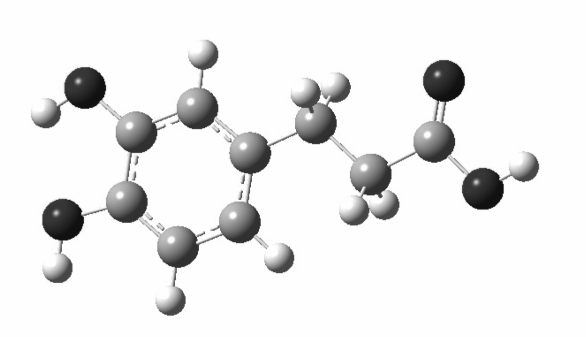


Fig. 3. Optimized structure of Dihydrocaffeic acid.

The most active scavengers in the group under consideration are Dihydrocaffeic acid (83.7 % RSA) and Rozmarinic acid (82.4 %RSA). Their HOMO energies (-0.222 eV и -0.221 eV) are higher than HOMO energies of other compounds like α-Tocopherol (-0.193 eV), Trolox (-0.198), TBHQ (-0.210) and BHA (-0.206).

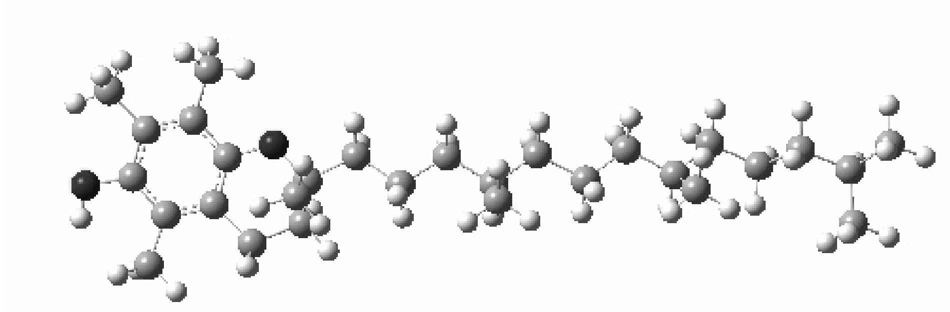
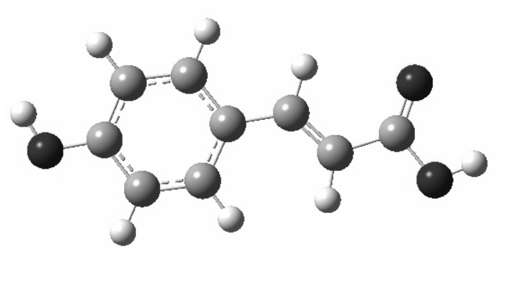


Fig. 4. Optimized structure of α-Tocopherol.

The scavenging activity of the latter compounds is 30 to 65 % lower than that of the former.

Dihydrocaffeic acid radical spin density (0.346) and C-O bond distance (1.260 Å) in are comparable to the spin density and C-O bond distance in the radicals of lower activity compounds - Chlorogenic acid (49.2 %RSA, C-O bond distance 1.253Å and spin density 0.287), and Sinapic acid (54.4 %RSA, C-O bond distance - 1.245Å and spin density 0.315).



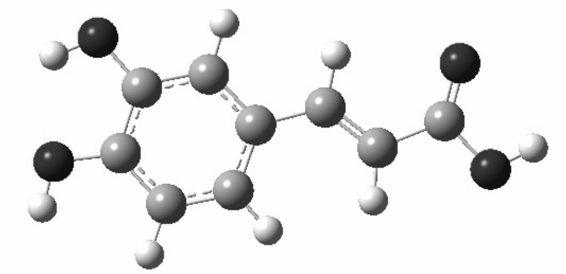


Fig. 5. Optimized structures of Coumaric acid and Caffeic acid.

The comparison of % RSA between *para*-Coumaric acid and Caffeic acid shows that the latter is by an order more active than the former. This difference can be due to the electronic influence of the second hydroxyl group upon the dissociable one as well as to the intra-molecular hydrogen bond. The spin density at O-atom in the radical of the Coumaric acid is lower (0.335) than that of the Caffeic acid (0.380). According to this index the former should be more active but the experimental values of the activities are reverse. More illustrative for the role of the second hydroxyl group is the comparison between activities of Tyrosol (2.6 %RSA) and Hydroxytyrosol (56.5 %). This difference is too big to be explained by insignificant differences in HOMO energies (0.006 eV) and in spin densities (0.060).

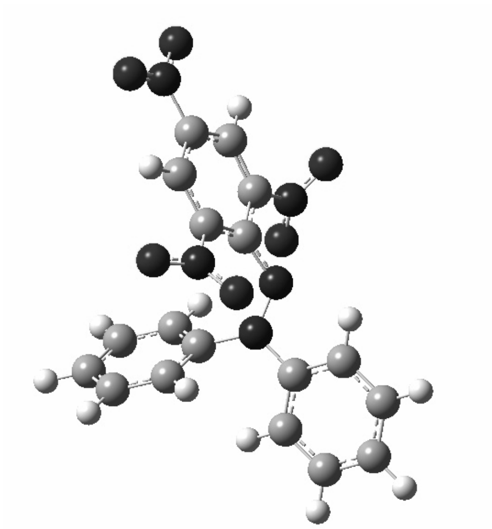


Fig. 6. Optimized structure of DPPH radical.

Another crucial factor for the DPPH-scavenging activity arises from the comparative analysis of the spatial models of the investigated compounds. It is the presence or absence of bulk substituent in the adjacent position to the dissociable hydroxyl group. All the compounds with adjacent methyl group are less active than the others which do not possess such a group. One of the least active compounds, BHT (5.7 %RSA), has HOMO energy comparable to the HOMO energy of the most active compounds. The indices of its radical are also close to those of the more active structures. α-Tocopherol and Trolox, which possess two methyl groups adjacent to the dissociable hydroxyl group and show medium activity (52.8 and 53.5%RSA) despite of their very low HOMO energies. We suppose that the reason lies in the particular spatial structure of DPPH-radical, namely two reactive centers are not able to reach each other as is shown in Fig. 6 and this is the rate determining factor.

4. CONCLUSIONS

In the present work the results of a preliminary analysis of the relation between DPPH-scavenging activity and some of the most popular descriptors of the antioxidant/scavenging activity of a series of natural antioxidants is presented. An absence of linear correlation was observed. The analysis of the obtained results shows that a correlation should be searched after appropriate dividing of the compounds according to: the presence of a second hydroxyl group; the existence of methyl groups in adjacent position. Such an analysis is planned for future investigations.

5. APPENDIX

The modified χ^2 criterion showed that the random variables X, Y and W are normally distributed. The population correlation coefficients ρ are defined as usual: for instance for variables X and Y by the eq. $\rho_{XY} = E(X - \mu_X)(Y - \mu_Y) / \sigma_X \sigma_Y$. We have tested the null hypotheses $H_0 : |\rho| = 0$ against the alternate hypotheses $H_1 : |\rho| \neq 0$. The critical value for sample size 15 at 95% confidence level is $r_{0,95} = 0.514$ [9]. The values of the sample correlation coefficients r_{YX} and r_{YW} are less than $r_{0,95}$ (cf. eq. 1 and eq. 3). Therefore the alternate hypotheses are rejected and the linear correlation between Y and X (resp. Y and W) is proved to be insignificant. The absence of significant linear correlation between variables Y and Z was proved by the use of Kenuy's criterion [9].

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PARTIAL AVERAGING FOR OPTIMAL CONTROL PROBLEMS WITH IMPULSIVE EFFECTS

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Abstract: *In this paper are present some results connected to the applications of the averaging method for solving of optimal control problems, where the models are systems of differential equations with impulsive effects. We suppose additional control in the impulses.*

Key words: *method of averaging, differential inclusion, impulsive differential inclusion, small parameter, controlled mechanical system.*

AMS subject classification: *49N25, 49J24, 49J25.*

1 INTRODUCTION

In works devoted to the application of the averaging method in optimal control, are developed two approaches:

1) Using necessary conditions for optimality, the original optimal control problem is redused to a boundary value problem, witch can be solve using the averaging method.

2) The equations of the moving object are averaged directly, and after thet one solves the optimal control problem ,for the redused system.

In this work we consider numerical-asymptotic methods for solving optimal control problems, witch are based on the second approach. Briefly, this approach includes the following:

1. To the nonautonomous optimal control problem, using various averaging schemes for differential inclusions (in our case differential inclusion with impulses) is assigned autonomous optimal control problem.

2. So obtained simpler optimal control problem can be solved by numerical analysis method, or analyticly.

This approach considerably reduse the calculations witch are necessary for solving of the original optimal control problem.

Let the differential equations of the movement of the controllable object be the following:

$$\dot{x} = \varepsilon\left[f(t,x) + f_1(t,x,u)\right], \ \ t \neq t_i, \ x(0) = x_0, \qquad (1)$$

$$\Delta x \,|_{t=t_i} = \varepsilon I_i(x, w_i), \qquad (2)$$

where $\varepsilon > 0$ is a small parameter, x is n-dimentional phase vector, f(t, x) and $f_1(t,x,u)$ are continuous and 2π - periodic with respect to t vector-functions, $u \in U \in comp(R^n)$ is a control of the system, and $w \in W \in comp(R^n)$ is a control by impulses. We suppose that $L_{i+p}(x,w) = I_i(x,w),\ t_{i+p} = t_i + 2\pi$, where p is integer.

We have to find admissible controls $\{u(t), w_i\}$, minimizing at the finite moment $t = T = L\varepsilon^{-1}(L = const)$ the functional of the following kind:

$$J[v, z_i] = \Phi(x(T)) \qquad (3)$$

For solving of this problem we can use some of the schemes for partially averaging.

Everywhere below we suppose that all lfunctions are uniformly bounded, and Lipshitz continuous with respect to x and u.

2. MAIN RESULT

Firs scheme for partially averaging.

To the control – problem (1) – (3) is assigned the following partially averaging problem

$$\dot{y} = \varepsilon\left[\bar{f}(y) + f_1(t,y,v)\right],\ t \neq t_i,\ y(0) = x_0, \qquad (4)$$

$$\Delta y\,|_{t=t_i} = \varepsilon I_i(y, z_i), \qquad z_i \in W \qquad (5)$$

$$J[v, z_i] = \Phi(y(T)) \to \min \qquad (6)$$

where

$$\bar{f}(y) = \frac{1}{2\pi}\int_0^{2\pi} f(t,y)dt, \qquad (7)$$

Second scheme for partially averaging. To the (1) – (3) we assigned the following partially averaged problem:

$$\dot{y} = \varepsilon\left[\bar{f}(y) + z(t) + f_1(t,y,v)\right],\ y(0) = x_0, \qquad (8)$$

$$J_2[v, z] = \Phi(y(T)) \to \min \qquad (9)$$

where $\bar{f}(y)$ is obtained by (7), z(t) is the new vector-control and

$$z(t) \in I_0(y) = \frac{1}{2\pi} \sum_{0<t\le 2\pi} I_i(y,W) \qquad (10)$$

The first case (4) – (7) is a optimal-control problem with controllable impulses, and optimal control problem (8) – (10) is without impulses.

The next theorem give proof of the utility of the first scheme of partial averaging, i. e. these theorems show, that for appropriate chosen ε the solutions of the system (1) – (3) are sufficiently close to the solutions of the system (4) – (7).

Let J^* and J_1^* be optimal values of the functionals (3) and (6) of the system (1) – (3) and of the averaging system (4) – (7), respectively. Let $J(v^*(t), z_i^*)$ be the value of the functional (3) obtained by the optimal control $v^*(t)$ and z_i^* of (4) – (7).

Theorem 1 Let in the domain Q, where

$Q = \{t \ge 0, x \in D \subset R^{n,}, u \in U \subset R^m, w_i \in W \subset R^m\}$

the following conditions be fulfilled:

1) the function $f(t,x)$ and $f_1(t,x,u)$ are piece-wice continuous and 2π - periodic with respect to t, satisfy Lipshitz condition w.r. to x with constant λ, and they are uniformly bounded with a constant M;
2) the function $f_1(t,x,u)$ is continuous w.r. to u and $f_1(t,x,U) \in conv(R^m)$;
3) the functions $I_i(x,w)$ are uniformly bounded by a constant M, satisfy Lipshitz condition w.r. to x with, constant λ and they are continuous w.r. to w.
4) $\forall \varepsilon \in (0,\sigma)$ $(\sigma = const)$ and all admissible controls $\{u(t), w_i\}$ the trajectories of the system (4), (5) with some ρ-neighbbourhood belong to D;
5) $\Phi(x)$ is Lipshitz function w.r. to x with a constant λ.

Then $\forall (L>0) \exists (C(L)>0) \exists (\varepsilon^0(L) \in (0,\sigma])$ such that if $\varepsilon \in (0,\varepsilon_0]$ the following estimations

$$\left| J^* - J_1^* \right| \le C\varepsilon \qquad (11)$$

$$J\left(v^*(t), z_i^*\right) - J^* \le C\varepsilon \qquad (11')$$

hold.

Proof. From the conditions of the theorem and from [5] (Theorem 2 p.265) it follows that the attainable sets of the systems (1), (2) and (4), (5) are compact sets and with some ρ-neighbbourhood belong to D. Therefore the problems (1) – (3) and (4) – (7) have optimal solution $\{J^*, x^*(t), u^*(t), w_i^*\}$ and $\{J_1^*, y^*(t), v^*(t), z_i^*\}$, respectively.

Let us write the system (1), (2) and (4), (5) as differential inclusions:

$$\dot{x} = \varepsilon\left[f(t,x) + f_1(t,x,U)\right], \; t \ne t_i, \; x(0) = x_0, \qquad (12)$$

$$\Delta x\,|_{t=t_i} \in \varepsilon I_i(x, W), \tag{13}$$

and

$$\dot{y} = \varepsilon\left[\overline{f}(y) + f_1(t, y, U)\right],\ t \neq t_i,\ y(0) = x_0, \tag{14}$$

$$\Delta y\,|_{t=t_i} = \varepsilon I_i(y, W), \tag{15}$$

It is obvious that differential inclusion (14), (15) is partialy averaged inclusion of the differential inclusion (12), (13) and for these system all conditions of Theorem 1 and remark 3 of [8] are fulfilled. Consequently for $\zeta > 0$ there exist $\varepsilon^0 = \varepsilon^0(\zeta)$ for which if y(t) is a solution of the system (14), (15), then there exist a solution x(t) of the system (12), (13) such that

$$\|x(t) - y(t)\| < \xi, if\, 0 < \varepsilon \leq \varepsilon^0$$

and vice versa.

When the right-hand side of the systems are $2\pi -$ periodical functions with respect to t, then there exist constants $\varepsilon > 0$ and C>0, such that for all $\varepsilon \in (0, \varepsilon^0]$ the above estimation has the following form ([10], [6])

$$\|x(t) - y(t)\| \leq C\varepsilon$$

Therefore $\exists(\varepsilon^0 \in [0,\sigma])\exists(c_1 > 0)\forall(\varepsilon \in (0, \varepsilon^0])$ the following estimation

$$h(X(T), Y(T)) \leq c_1\varepsilon \tag{16}$$

hold,

where X(T) and Y(T) are attainable sets of the differential inclusion (12) – (13) and (14) – (15) respectively or attainable sets of the system (1) – (2) and (3)- (4) ($c_1 = const$).

Obviously

$$J^* = \min_{x \in X(T)} \Phi(x) \qquad J_1^* = \min_{y \in Y(T)} \Phi(y) \tag{17}$$

According to (16), (17) and condition 5) of the theorem (Lipshitz condition for $\Phi(x)$ w.r. to x) we are going to estimate

$$\left|J^* - J_1^*\right| \leq \left|\min_{x \in X(T)} \Phi(x) - \min_{y \in Y(T)} \Phi(y)\right|$$

Let $\Phi(x^*) = \min_{x \in X(T)} \Phi(x)$ ($x^*(t)$ is optimal solution of (1) – (3)).

Let $\Phi(y^*) = \min_{y \in Y(T)} \Phi(y)$ ($y^*(t)$ is optimal solution of (4) – (7)).

Since $h(X(T), Y(T)) \leq c_1\varepsilon$, then for $y^* \in Y(T)$ there exist $\overline{x} \in X(T)$, such that $\|y^* - \overline{x}\| \leq c_1\varepsilon$, and for $x^* \in X(T)$ there exist $\overline{y} \in Y(T)$, such that $\|x^* - \overline{y}\| \leq c_1\varepsilon$.

From the fact that $\Phi(.)$ is Lipshitz function and from the above estimation we have

$$\left|J_1^* - J(\overline{x})\right| = \left|\Phi(y^*) - \Phi(\overline{x})\right| \leq \lambda c_1 \varepsilon \tag{λ1}$$

$$\left|J^* - J_1(\bar{y})\right| = \left|\Phi(x^*) - \Phi(\bar{y})\right| \le \lambda c_1 \varepsilon \qquad (\lambda 2)$$

Obviously $J^* \le J(\bar{x})$ and $J_1^* \le J_1(\bar{y})$

For J^* and J_1^* there are two possibilities:

a) $J^* \le J_1^* \Rightarrow J^* \le J_1^* \le J_1(\bar{y})$ and according to ($\lambda 2$)

$$J_1^* - J^* \le J_1(\bar{y}) - J^* \le \lambda c_1 \varepsilon \Rightarrow \left|J^* - J_1^*\right| \le \lambda c_1 \varepsilon$$

b) $J_1^* \le J^* \Rightarrow J_1^* \le J^* \le J_1(\bar{x})$ and according to ($\lambda 1$)

$$J^* - J_1^* \le J(\bar{x}) - J_1^* \le \lambda c_1 \varepsilon \Rightarrow \left|J_1^* - J^*\right| \le \lambda c_1 \varepsilon.$$

Substituting $C_1 = \lambda c_1$ and from a) and b) we obtain (11) , so

$$\left|J^* - J_1^*\right| = C_1 \varepsilon \qquad (18)$$

Now let us prove inequality (11').

We consider the equations

$$\dot{x}^1 = \varepsilon\left[f(t,x^1) + f_1(t,x^1,v^*(t))\right], \;\; t \ne t_i, \; x^1(0) = x_0, \qquad (19)$$

$$\Delta x^1 \big|_{t=t_i} = \varepsilon I_i(x^1, z_i^*), \qquad (20)$$

which are obtained from the system (1), (2) by the substitutions $u(t) = v^*(t)$ and $w_i = z_i^*$.

The controls $v^*(t)$ and z_i^* are obtained by solving the system (4), (5).

We also consider the equations

$$\dot{y}^* = \varepsilon\left[\bar{f}(y^*) + f_1(t,y^*,v^*(t))\right], \;\; t \ne t_i, \;\; y^*(0) = x_0, \qquad (21)$$

$$\Delta y^* \big|_{t=t_i} = \varepsilon I_i(y^*, z_i^*), \qquad (22)$$

The equations (21), (22) are partially averaged with respect to the equations (19), (20).

Therefore $\exists(\varepsilon^0 \in (0,\sigma])\exists(C_2 > 0)\forall(\varepsilon \in (0,\varepsilon^0])\forall(t \in (0, L\varepsilon^{-1}[)$ the following estimation

$\left\|x^1(t) - y^*(t)\right\| \le C_2 \varepsilon$ hold.

Then

$$\left|J_1^* - J\left[v^*(t), z_i^*\right]\right| = \left|\Phi(y^*(T)) - \Phi(x^1(T))\right| \le \lambda \left\|x^1(T) - y^*(T)\right\| \le \lambda C_2 \varepsilon$$

i.e

$$\left|J_1^* - J\left[v^*(t), z_i^*\right]\right| = \le \lambda C_2 \varepsilon \qquad (23)$$

According to (18) and (23) we have:

$$J\left[v^*(t), z_i^*\right] - J^* = \left|J^* - J\left[v^*(t), z_i^*\right]\right| == \left|J_1^* - J\left[v^*(t), z_i^*\right] + J^* - J_1^*\right| \le$$

$$\le \left|J^* - J_1^*\right| + \left|J_1^* - J\left[v^*(t), z_i^*\right]\right| \le C_1 \varepsilon + \lambda C_2 \varepsilon$$

Substituting $C = C_1 + \lambda C_2$ we complete the proof. **Q.E.D.**

Remark 1. From Theorem 1 is follows, that the optimal solution of the partially averaged system (with respect to the value of the functional), is differ from the exact solution of the original system no more than $C\varepsilon$. Except this, when one use the optimal control of the averaged problem as a control of the original problem, one obtain a solution which is different from the exact solution (with respect to the functional) but no more than $C\varepsilon$.

The next theorem is corollary of Theorem 1, but we separate it because it is related to time optimal control problem.

Theorem 2. Let the conditions 1)- 4) of Theorem 1 be fulfilled and $\{y^*(t), v^*(t), z_i^*, T^*\}$ is a solution of the time optimal control problem about optimal reduction of the system (4), (5) to a given last state $y(T^*) = b$.

Then $\exists(\varepsilon^0 \in (0,\sigma])\exists(C>0)\forall(\varepsilon \in (0,\varepsilon^0])$ the controls $u(t) = v^*(t)$ and $w_i = z_i^*$ guarantee the redusing of the system (1), (2) into $C\varepsilon$ - neighbourhood of the point b for time no more than T^*.

Proof. From the proof of the Theorem 1 follow

$$\|y^*(T^*) - x^1(T^*)\| \le C\varepsilon \qquad (24)$$

Here $x^1(t)$ is a solution of (19), (20).

Since $y^*(T^*) = b$ then $\|x^1(T^*) - b\| \le C\varepsilon$.

By analogy with Theorem 1 one can prove similar theorem for the second scheme.

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EDUCATION ON COMPUTER NETWORKS IN SOUTH- WEST UNIVERSITY

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***Abstract**: The aim of the present paper is to propose some labs with the network equipment of Department of Computer Science and Technologies which includes three Bridges D-Link DI-1100i, one CISCO Switch WS-C2950-12, three Routers CISCO 831-K9-64; Ethernet network with 12 workstations, and a MS Windows 2003 server, Internet, Protocol analysis software.*
***Keywords:** Computer Networks, Dispersion analysis, and Education.*

1. INTRODUCTION

The aim of the course of Computer Networks is to give the students knowledge and skills on the basic principles, standards and tendencies of development in the field of computer networks. This will help them in future to professionally solve system tasks in the area of computer networks and communications. This course is compulsory for the last year students of specialty "Informatics" – bachelor degree. Basic knowledge in computers and informatics are prerequisite.

Teaching methods include lectures- 3 hours per week or total 45 hours, as well as laboratory work (based on instructions, experiments and draw up protocols) – 10 Labs X 3 hours = 30 hours.

The Network Equipment of the Department of Computer Science and Technologies for laboratory work includes: 3 Bridges D-Link DI-1100i, one Switch WS-C2950-12 and 3 Routers CISCO 831-K9-64; Ethernet network with 12 workstations, and a MS Windows 2003 server, Internet, Protocol analysis software, and tutorial for every lab.

The aim of the present paper is to propose some labs introduced for education on "Computer Networks" with the network equipment mentioned above.

2. COMPUTER NETWORKS LABS

The purpose of Lab 1 is to acquaint students with the basic peripheral components of a PC computer system and their connections including network attachment. Students examine the internal PC configuration and identify major components, too. Students observe the boot process for the Windows operating system and use the Control panel to find out information about the PC. Knowing the components of a PC is valuable when troubleshooting and is important to success in the networking field. The Lab includes realization and investigation of communication between PCs via their serial interfaces- RS232C [1, 2]. Since the serial interfaces on the (every couple) workstations are directly connected students will not be able to connect any additional workstations.

The purpose of the Lab 2 is to acquaint students with the network settings required to connect PC to a local area network and to gain access to the Internet (World Wide Web - WWW) and Intranet (internal local web servers). Students review Network Interface Card

(NIC) configuration, drivers, and TCP/IP protocol settings for a typical Windows client workstation in a server based Ethernet network. In this lab students learn how to use the workstation network settings when must set up workstations or have a problem logging onto a network.

The purpose of Lab 3 is students to build and test Unshielded Twisted Pair (UTP) Category 5 Ethernet and fiber optic network cables.

The Labs 4 and 5 help students develop an understanding of IP addresses and how TCP/IP networks operate, investigate Internet Control Message Protocol (ICMP), as well as subneting.

In Lab 6 students learn how to build Virtual LANs (VLANs) on the switch- WS-C2950-12 using http facility or console commands two PCs to create a simple Peer-to-Peer LAN or workgroup.

In lab 7 students focus on Spanning-Tree Protocol- STP. Companies are increasingly looking for 24 hour, seven day a week uptime for their computer networks (achieving over 99.999% uptime). A network that is based on switches or bridges will introduce redundant links between those switches or bridges to overcome the failure of a single link. These connections introduce physical loops into the network. A physical topology that contains switching or bridging loops is necessary for reliability, yet a switched network cannot have loops. A redundant switched topology may cause broadcast storms, multiple frame copies, and MAC address table instability problems. If there will be Broadcast Storm, the network will appear to be down or extremely slow. The solution is to allow physical loops, but create a loop free logical topology, and is called a tree. This topology is a star or extended star logical topology, the spanning tree of the network. It is a spanning tree because all devices in the network are spanned.

The purpose of this Lab is students to create a basic switch configuration and verify it, determine which switch is selected as the root switch with the factory default settings, force the other switch to be selected as the root switch, and observe the behavior of spanning-tree algorithm in presence of switched network topology changes.

The Spanning-Tree Protocol requires network devices to exchange messages to detect bridging loops. Links that will cause a loop are put into a blocking state. The message that a switch sends, allowing the formation of a loop free logical topology, is called a Bridge Protocol Data Unit- BPDU (fig. 1).

STP Algorithm. Spanning Tree Protocol uses the spanning-tree algorithm to construct a loop free shortest path network. The algorithm is used to:

(i) Select a single switch that will act as the root of the spanning tree.

(ii) Calculate the shortest path from itself to the root switch. Shortest path is based on cumulative link costs. Link costs are based on the speed of the link.

(iii) Designate one of the switches as the closest one to the root, for each LAN segment. This bridge is called the “designated switch”. The designated switch handles all communication from that LAN towards the root bridge.

(iv) For each non root switch, choose one of its ports as its root port. This is the interface that gives the best path to the root switch (has a lower Root Path Cost to the Root Bridge than its other ports).

(v) Select ports that are part of the spanning tree, the designated ports. Non-designated ports are blocked. The STA places all Root Ports and Designated Ports into forwarding state, while the others into blocking state. In the blocking state, ports can only receive BPDUs, Data frames are discarded and no addresses can be learned.

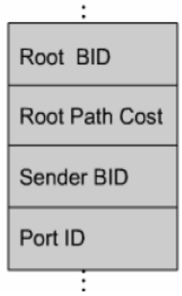

Fig. 1: Format of BPDU: Root Path Cost shows how far away is the Root Bridge.

The BIDs consist of a bridge priority (that defaults to 32768) and its base MAC address. Root BID holds ID of Root Bridge. Sender BID identifies the Sender Bridge. Port ID shows which port sends this BPDU.

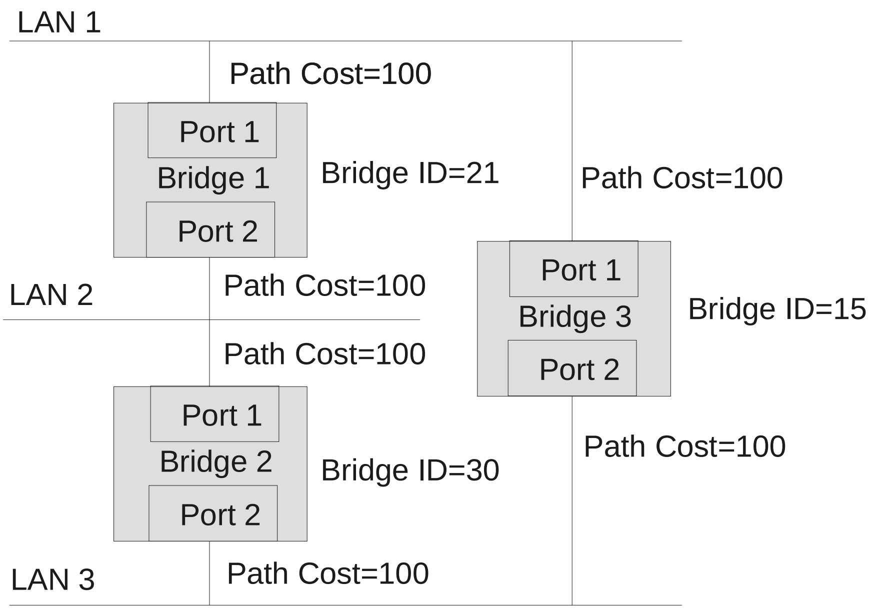

Fig. 2. Network example.

Selection of Root Bridge. When a switch first starts up, it assumes it is the root switch and sends BPDUs which contain the switch BID in both the root and sender Bridge ID (BID). BPDUs are sent out with the BID. As a switch receives a BPDU with a lower root BID it replaces that in the BPDUs that are sent out. All bridges see these and decide that the bridge with the smallest BID value will be the root bridge.

In the following example (fig. 2), three bridges are used to connect three LANs together.

The results of STP are: One root bridge per network- Bridge 3; One root port per non root bridge- Port 1 of Bridge 1 and Port 2 of Bridge 2; One designated port per segment- both ports of Bridge 3 and Port 2 of Bridge 1; Unused, non-designated ports- Port 1 of Bridge 2.

Bridge 3, as the Root Bridge, also becomes the Designated Bridge for LAN 1 and LAN 3. Bridge 1 becomes the Designated Bridge for LAN 2. Only Designated Bridges have Designated Ports, so Bridge 2 does not have any Designated Port. For Bridge 1, Port 1 is the Root Port , while Port 2 is the Designated Port, because Port 1 of Bridge 1 has a lower Root Path Cost to the Root Bridge than its Port 2.

Laboratory exercises:

1. Cable a network to the diagram (fig. 4), create a basic switch configuration and verify it..

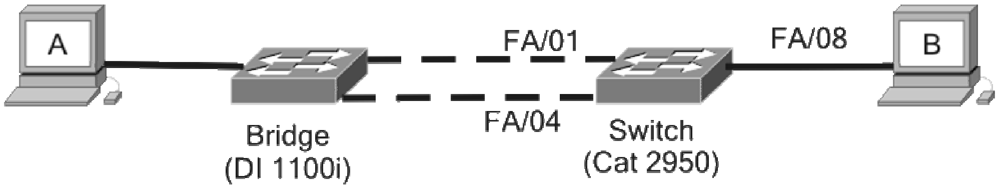

Fig. 4. Network topology.

2. Determine that the switch selected as the root bridge. Examine MAC addresses of Bridge and Root Bridge in the dialog box of DI 1100i SNMP management software (fig. 5).

3. Display the switch spanning tree table as type follows:

Switch_B#show spanning-tree brief

Examine the output and answer the following questions: Which switch is the root switch? What is the priority of the root switch? Which ports are forwarding on the root switch? Which ports are blocking on the root switch? What is the priority of the non-root switch?

4. Reassign the root bridge. Force the other device to become the root switch, by changing default values for priority (from 32768 to 1024). It is necessary to set the priority according to fig. 5, or at the privileged EXEC mode prompt, enter the following:

Switch_B(config)#spanning-tree vlan 1 priority 1024
Switch_B(config)#exit

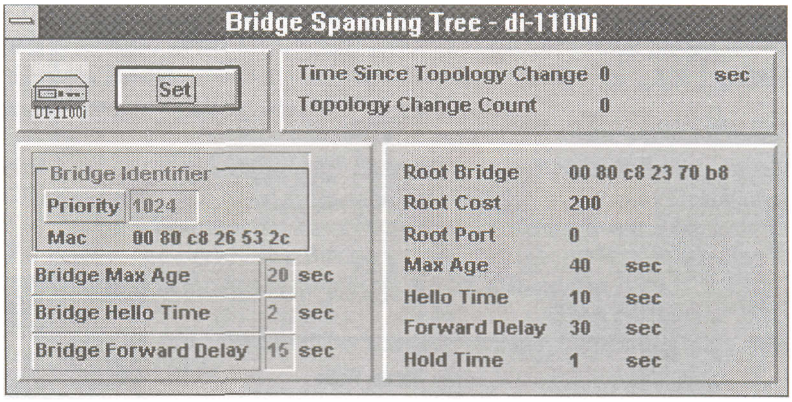

Fig. 5. The Bridge Spanning Tree dialog box of DI 1100i
SNMP management software

5. Look at the spanning tree table on switch and answer the questions from step 3.

6. Remove the cable from the forwarding port on the non-root switch. Wait for at least two minutes. Observe the behavior of spanning-tree algorithm in presence of switched network topology changes.

7. Display the switch spanning tree table. On Switch_B type show spanning-tree brief at the privileged EXEC mode prompt and fill in a table. What has happened to the switch port LEDs?

8. Replace the cable in the port that it was removed from and repeat the experiment.

The rest of labs focus on routed and routing protocols[2] , Directory and Domain Name System (DNS) services in OS Windows Server 2003, as well as issue of troubleshooting.

3. CONCLUSIONS

In present paper have been proposed some labs for education on the subject “Computer Networks” with the network equipment which includes three Bridges D-Link DI-1100i, one CISCO Switch WS-C2950-12, three Routers CISCO 831-K9-64; Ethernet network with 12 workstations, and a MS Windows 2003 server, Internet, Protocol analysis software. The labs were introduced for education on “Computer Networks” in 2006/7 year.

In order to evaluate how much valuable results at learning have been achieved using these labs I will apply dispersion analysis of the examination results of students in 2005/6 and 2006/7.

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WEBMONITOR – WEB BASED DATA ACQUISITION SYSTEM FOR TEMPERATURE MEASUREMENTS

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Abstract: *This paper presented a solution for the incorporation of graphs & charts into web/intranet pages and applications. Versatile components provide the ability for web authors and Java developers to easily build and publish dynamic and interactive graphs & charts. With the advanced graphing functionality you will be quickly adding impressive and dynamic charting capabilities bringing your data alive. The PHP graphing scripts provide a very easy way to embed dynamically generated graphs and charts into PHP applications and HTML web pages. The graphing software is very easy to use and it is perfectly possible to add professional quality real time graphing to web pages / applications within minutes.*

Keywords: *Web monitoring, scripts, software.*

1. INTRODUCTION

Web based data acquisition (or web ready data acquisition) is another term for networked data acquisition and control. Networked data acquisition and control is technology that allows user's to use computer networks or the Internet to directly access and monitor processes and systems, acquire application data, and maintain control points. This Webmonitor enables dynamic real-time graphing functionality to be added to any type of web page, including HTML documents, PHP, JSP, ASP etc.

2. OVERVIEW

Presented in this paper web based data acquisition system for temperature measurements is developed with microcontrollers from family PIC. PIC'S are general purpose micro controllers, manufactured by Microchip Technology. They are popular devices for hobbyists/ nerds/ electronic geniuses and can be used for all sorts of purposes. PIC's come in different packages, speeds, voltage range, temperature range and power dissipation. The 16F84-04/P is the "entry level" version. Its siblings will probably work in the circuit, but I haven't tested all of them. A 16F84-04/P device needs programming before it can do anything useful. The PIC Source .asm file needs to be compiled. The generated .hex file will contain the program instructions and configuration bits. The .hex file is read by the programmer and is used to "blow" the device. A 16F84-04/P can be reprogrammed many times.

The temperature sensors are DS18S20 of Dallas Semiconductor. This sensors used 1-Wire® protocol how was originally designed for communication with nearby devices on a short connection — a way to add auxiliary memory on a single microprocessor port pin. Customers soon devised unique applications that involved extending the bus and moving the slave devices farther and farther from the master. Problems came up as the bus lengths exceeded both the capabilities of the bus masters and the limits of the protocol. 1-Wire device designs responded to the call with added features and protocols, multi-drop (networking) capabilities, durable steel containers (iButtons®), and mechanisms to assure valid data transfers even in severely intermittent contact situations.

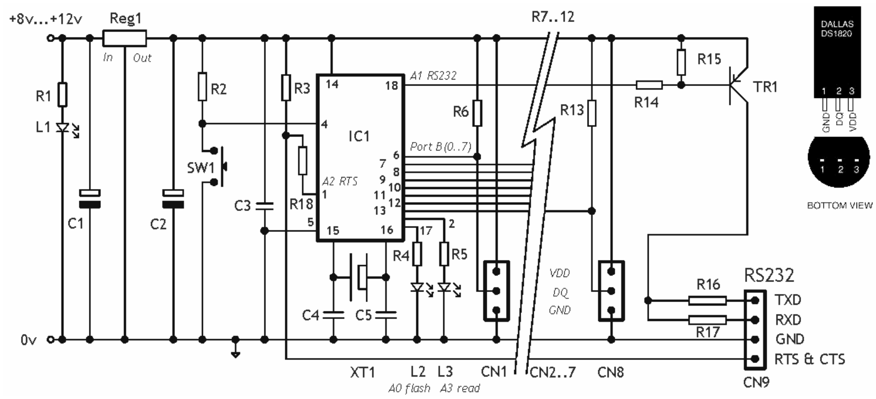

Fig.1.

Schematic of the data acquisition system

For communication between PC and board with microcontroller was developing the windows application with visual language Delphi (fig.2.). This application read the data for temperature from microcontroller and transferred them in format for web presentation. Data base with record in file is too available.

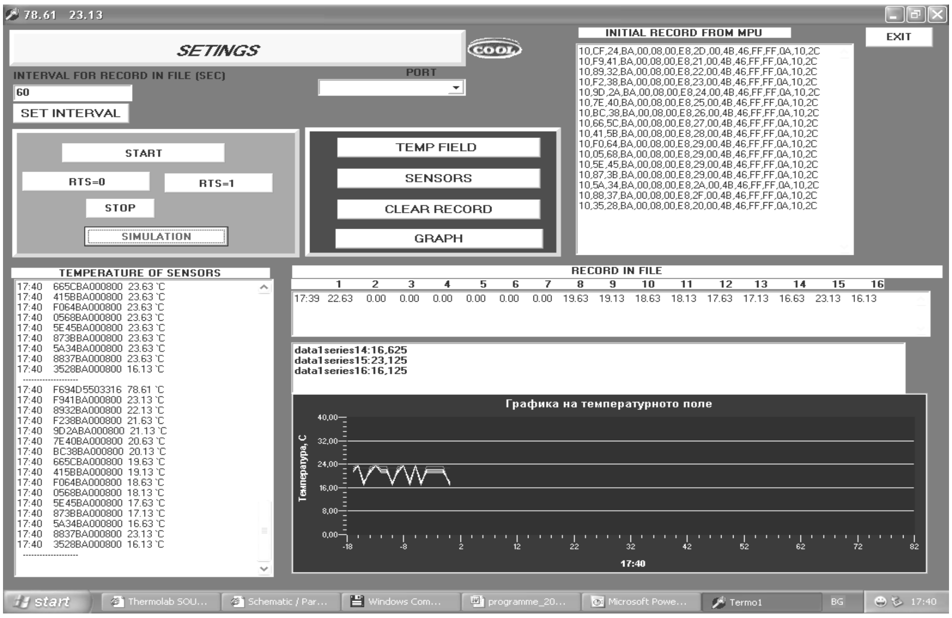

Fig.2. PC application for data acquisition

3. WEBMONITOR

The Webmonitor package is a piece of software written entirely within the PHP language. The software can operate on any web server with a PHP Engine. The set-up and install process is very easy and simply requires copying a few files to your web server.

A graph is added to a web page using the standard html <IMG> tag (the same tag which is used to add images to a web page). In this case the <IMG> tag is set to point at the graphing software, rather than an actual image file.

When a user's browser views a page which contains the graphing <IMG> tag the following will occur (automatically).

1. The browser will send a request to the Webmonitor software for the graph image.
2. Upon receiving a request the Webmonitor software will load up the configuration file and acquire the data from the source specified.
3. The graphing software will then create a graph image according to the configuration and data loaded.
4. The graph image is then formatted in standard PNG format and returned to the requesting browser.
5. The final step is the browser displaying the graph image to the user.

All of this happens automatically and very quickly, in fact the user will not notice the difference between this and any other image within the page.

Unlike some graphing software packages, at no point is the graph image stored as a file on the web server. The graph image is always created 'on the fly' and all system resources are immediately released. This makes the graphing package incredibly scalable and easy to manage.

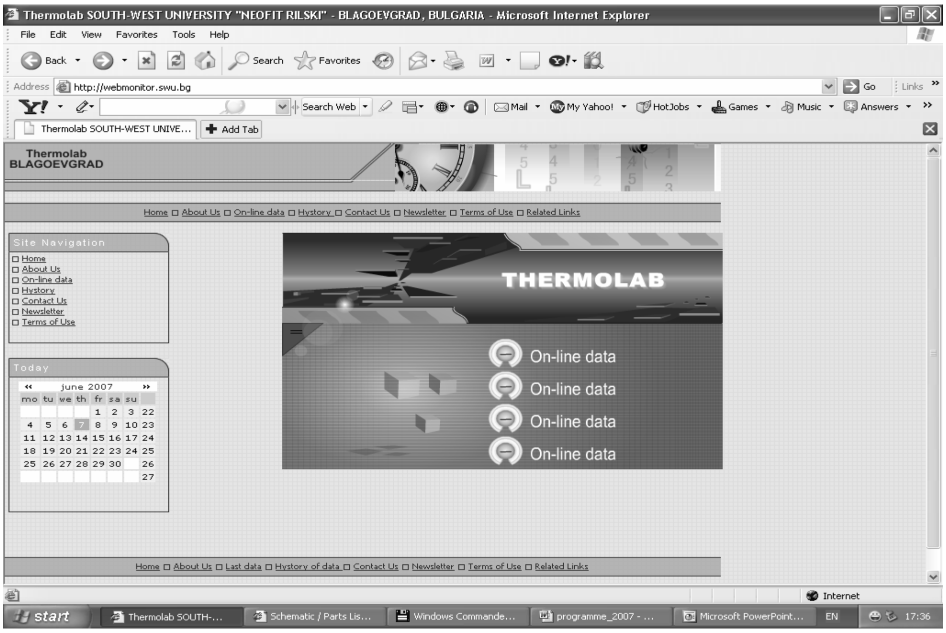

Fig.3. Web page webmonitor.swu.bg

Installing the Webmonitor on your web server is very easy. The contents of the Webmonitor directory can actually be placed within any of your web directories. You have now installed the graphing software – Webmonitor. You should now be able to access and see the graphing software working by entering the following URL in your browser's address bar:

http://www.yourdomain.com/webmonitor/index.htm

Requirements are Web Server running PHP 4.0.6 or higher. Most web servers do now include a PHP engine, but if you are unsure then either ask your server administrator/hosting provider or just give it a try anyway.

3.1. Retrieving Data from files

Data may be added to the graph in one of four ways:

- Retrieving Data from files
- Retrieving Data from another server process
- Retrieving Data from a MySQL Database
- A PHP Data Script

In this is present first option. To set the graph to read the data from file simply add the "data" parameter to the URL string of the <IMG> tag. For example if the data file is "graphdata.txt" then your <IMG> tag would become:

```
<img src="line-graph.php?data=graphdata.txt"
width=500 height=500>
```

Within the file the data should follow this format:

dataNseriesM: [value]

where N represents the position of the data item in the series and M represents the series number. For example for 1 series of data each containing 6 points the contents of the data file would be:

```
data1series1: 30
data2series1: 20
data3series1: -10
data4series1: 40
data5series1: 50
data6series1: 60
```

3.2. Parameter Reference

There are many configurable options for the Webmonitor. This section describes some parameter and it's effect (fig.4.). With the exception of 2 parameters, all are optional and if not supplied the graph will automatically calculate the values. The only 2 parameters which must be supplied in all cases are width and height.

General Graph properties

Parameter Name	Range of Values	Example
width	Specifies the width in pixels	width: 300
height	Specifies the height in pixels	height: 400
ndecplaces	Specifies the number of decimal places to use when displaying values	ndecplaces: 2
thousandseparator	Any character symbol.	thousandseparator: ,
backgroundcolor	general background color of the graph image	backgroundcolor: white
connectinglines	determines whether points should be connected or not	connectinglines: true
displayvalues	specifies whether actual values	displayvalues: false

	should be displayed at the top of each point	
bgimage	URL to an Image file specifies an image to used as the background	bgimage: brimage.gif

Grid Properties

Parameter Name	Range of Values	Example
grid	specifies whether to draw grid lines or not	grid: true
axis	specifies whether to draw axis lines or not	axis:
nrows	number of grid rows	nrows: 10
hspace	space in pixels of each grid column	hspace: 30
vspace	space in pixels of each grid row	vspace: 30
gridstyle	specifies the line style of the grid lines	gridstyle: dotted
gridcolor	Grid line color	gridcolor: #888888
axiscolor	Axis line color	axiscolor: #000000
floorcolor	If 3D is on then this specifies the color of the x-axis floor	floorcolor: #555599
gridbgimage	specified an image to used as the grid background	gridbgimage: brimage.gif
gridbgcolor	the background color of the grid area	gridbgcolor: light grey
gridposition	specifies the position of the bottom left of the grid	gridposition: 30,275
gridlineh	specifies whether to draw horizontal grid lines or not	gridlineh: false
gridlinev	specifies whether to draw vertical grid lines or not	gridlinev: false
gridbgcolor2	specifies a second background color of the grid area. If this parameter is specified then grid rows are colored alternating between this color and gridbgcolor	gridbgcolor2: grey

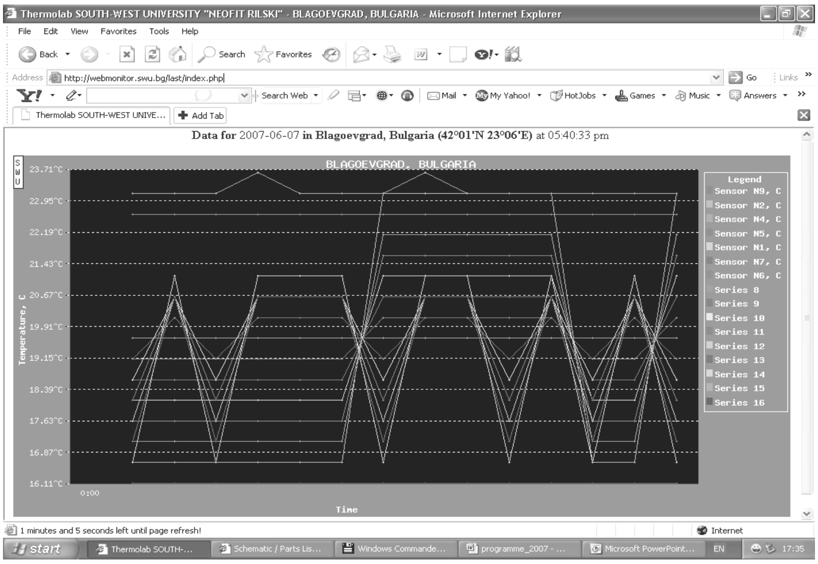

Fig.4. Web interface of the data acquisition system

Other parameters are optional:

- Scale properties
- 3D properties
- Legend properties
- X Axis Labels
- Y Axis Labels
- Graph Titles
- Series Specifications
- Target Lines
- Trend Line
- Free Form Text
- Free Form Images

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ANALYSIS OF ASSESSMENT RESULTS ON COMPUTER NETWORKS

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***Abstract**: The aim of the present paper is to describe the planning and conducting of an experiment within the students' groups educated in 2005/6 and 2006/7 years on "Computer Networks". The aim is to compare how different laboratory work exert influence on the students' learning process. Evaluation criteria are defined for theoretical, practical skills and expected results. The obtained results were statistical analyzed.*

***Keywords:** Computer Networks, Dispersion analysis, and Education.*

1. INTRODUCTION

The course of Computer Networks discuses the problems concerning design, building and application of computer networks[2]. The lectures include introduction to computer networks, principles of building, their contemporary classification, and architecture. ISO seven layers model of the Open System Interconnection- OSI, and Transmission Control Protocol/Internet Protocol - TCP/IP model are presented. The Internet is based on TCP/IP, which has become the standard "language" of networking. Although the TCP/IP model is the most widely used the seven layers of the OSI model are the ones most commonly used to describe and compare networking software and hardware from various vendors. It is very important to know both the OSI and TCP/IP models and be able to relate (or map) the layers of one to the other. Teaching course includes basic principles of building and functioning of Local Area Networks (LAN) illustrated by practical technical solutions in Ethernet and wireless LAN. The lectures on the most popular in the world computer network Internet present its basic characteristics, principles of functioning and application. The labs help students develop an understanding lecture material as well as contribute to formation of their skills.

Teaching methods include lectures- 3 hours per week or total 45 hours, as well as laboratory work (based on instructions, experiments and draw up protocols) – 10 Labs X 3 hours = 30 hours.

The Network Equipment of the Department of Computer Science and Technologies for laboratory work includes: 3 Bridges D-Link DI-1100i, one Switch WS-C2950-12 and 3 Routers CISCO 831-K9-64; Ethernet network with 12 workstations, and a MS Windows 2003 server, Internet, Protocol analysis software, and tutorial for every lab. Some labs with this network equipment were introduced for education on "Computer Networks" in 2006/7 year.

The aim of the present paper is to evaluate how much valuable results at learning have been achieved using these labs, as apply dispersion analysis of the examination results of students in 2005/6 and 2006/7.

2. COMPUTER NETWORKS LABS

The purpose of Lab 1 is to acquaint students with the basic peripheral components of a PC computer system and their connections including network attachment. Students examine the internal PC configuration and identify major components, too. Students observe the boot process for the Windows operating system and use the Control panel to find out information about the PC. Knowing the components of a PC is valuable when troubleshooting and is important to success in the networking field. The Lab includes realization and investigation of communication between PCs via their serial interfaces- RS232C [1, 2]. Since the serial interfaces on the (every couple) workstations are directly connected students will not be able to connect any additional workstations.

The purpose of the Lab 2 is to acquaint students with the network settings required to connect PC to a local area network and to gain access to the Internet (World Wide Web - WWW) and Intranet (internal local web servers). Students review Network Interface Card (NIC) configuration, drivers, and TCP/IP protocol settings for a typical Windows client workstation in a server based Ethernet network. In this lab students learn how to use the workstation network settings when must set up workstations or have a problem logging onto a network.

The purpose of Lab 3 is students to build and test Unshielded Twisted Pair (UTP) Category 5 Ethernet network cables. The students test straight-through , and crossover cables for good connections (continuity) and correct pinouts- correct color of wire on the right pin. The straight through cable means that the color of wire on pin 1 on one end of the cable will be the same as pin 1 on the other end. Pin 2 will be the same as pin 2 and so on (for the all eight wires). It will be wired to TIA/EIA-568-B or A standards for Ethernet which determines what color wire is on each pin. The specification T568-B is more common, but many installations are also wired to T568-A. In this lab students learn the wire mapping features of Cable Tester.

The Lab 4 helps students develop an understanding of IP addresses and how TCP/IP networks operate. IP addresses are used to uniquely identify individual TCP/IP networks and hosts (workstations and servers) on networks in order for devices to communicate. Hosts on a TCP/IP network in the beginning check if they have a unique IP address (which is referred to as its host address) as they send ICMP request and wait for replay or timeout. In this lab students investigate Internet Control Message Protocol (ICMP).

The Lab 5 helps students understand the basics of IP subnet masks and their use with TCP/IP networks. The subnet mask can be used to split up an existing network into subnetworks. This may be done to: i) reduce the size of the broadcast domains (create smaller networks with less traffic), ii) to allow LANs in different geographical locations to communicate or iii) for security reasons to separate one LAN from another. This Lab reviews the Default Subnet Mask and then focus on Custom Subnet Masks which will use more bits than the default subnet mask by "borrowing" these bits from the host portion of the IP address. This creates a three-part address; i) The original network address assigned, ii) The subnet address made up of the bits borrowed and iii) the host address made up of the bits left after borrowing some for subnets.

In Lab 6 students learn how to connect two PCs to create a simple Peer-to-Peer LAN or workgroup. The instructions for this lab focus on the Windows operating system. Students share a folder on one workstation and/or MSWindows 2003 server, and connect to that folder from the other workstation. This lab is divided into 2 exercises as follows:

- The two PCs (or workstations) are connected directly to each other from one Network Interface card (NIC) to the other NIC using a crossover cable. This can be useful to allow students to create a minilab for testing purposes without the need for a hub/switch.

- The two PCs are connected with a hub/switch between them. Thus, it allows for more than just two workstations to be connected depending on the number of ports on the switch. Using http facility or console commands the students build Virtual LANs (VLANs) on the switch- WS-C2950-12. The VLANs can have anywhere from two to twelve ports.

In lab 7 students focus on Spanning-Tree Protocol- STP. The purpose of this Lab is students to create a basic switch configuration and verify it, determine which switch is selected as the root switch with the factory default settings, force the other switch to be selected as the root switch, and observe the behavior of spanning-tree algorithm in presence of switched network topology changes.

The Lab 8 build on network topology and help develop a better understanding of IP routed and routing protocols [1,2] using a real-world example This exercise is based on foundations established in the prior labs. This lab focuses on a Class C networks, and using three Routers which separate these networks. Each router determines when a packet can go from one network to another.

The rest of labs help students develop a better understanding of the Directory and Domain Name System (DNS) services in OS Windows Server 2003, as well as issue of troubleshooting.

3. ANALYSIS OF ASSESSMENT RESULTS

The dispersion analysis [3, 4] made on the assessment results from the examination on the subject "Computer Networks" is given bellow. Two extracts are made for each group and the received results are given in histograms (fig.6 and fig.7 depict assessment results in 2005/6 and 2006/7). The proposed labs were introduced for education on "Computer Networks" in 2006/7 year.

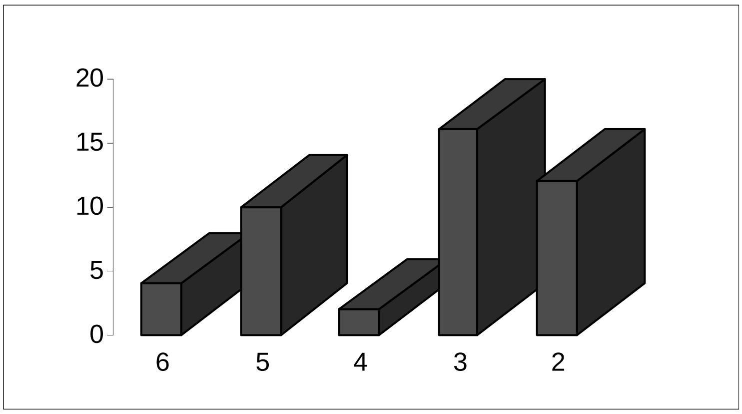

Fig.6 Assessment results in 2005/6

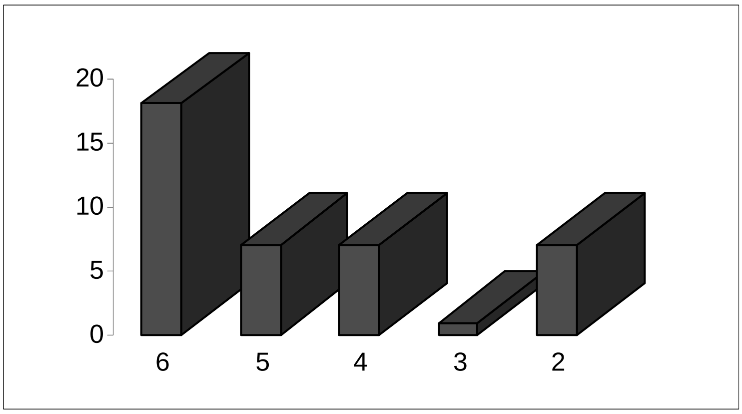

Fig.7 Assessment results in 2006/7

The course of the dispersion analysis is shown in Table 2. The general deviation is calculated by a consecutive subtraction of the general arithmetic mean mark ($\bar{x} =$ 4.07) from the examination mark of every student and the obtained results are squared:

(1) $\sum (x - \bar{x})^2 = 197.571$

Table2. The course of dispersion analysis.

Year	Number of students (f_i)	Sum of points from the test (x)	Arithmetic mean of the different groups ($\bar{x}_i$)	$\bar{x}_i - \bar{x}$	$(\bar{x}_i - \bar{x})^2$	$(\bar{x}_i - \bar{x})^2 \; f_i$
2005/6	44	154	3.5	-0.57143	0.32653	14.36735
2006/7	40	188	4.7	0.62857	0.3951	15.80408
Total	84	342	4.07			30.17143

The deviation calculation between the groups is given in Table 2 (see right side). The deviation inside the groups, or the sum of squares from the differences between the examination marks of individual students in the groups and the group arithmetic means, is found by the general and between the groups deviations [3] that already have been calculated:

(2) $\sum (\bar{x} - x_i)^2 = \sum (x - \bar{x})^2 - \sum (x_i - x)^2 \; f_i = 197.571 - 30.171 = 167.4$

The variation degrees of freedom between the groups are k – 1 = 1 since the groups are two (k = 2) in general, and inside the groups variation degrees are respectively n - k = 82 (n = 84).

The two dispersion estimations are found on the basis of deviations between the groups and inside the groups:

(3) $\sigma_M^2 = \dfrac{\sum (\bar{x}_i - x)^2 \; f_i}{k-1} = 30.171$

(4) $\sigma_B^2 = \dfrac{\sum (x - \bar{x}_i)^2}{n-k} = 2.041$

The ratio between the two assessments of the dispersion is:

(5) $F = \dfrac{\sigma_M^2}{\sigma_B^2} = 14.779$

In order to verify whether the proposed above labs considerably exerts influence on the quality of education, a level of significance 0.01 is chosen. The critical value from F distribution at significance level of 0.01, and degrees of freedom k – 1 = 1 and n – k = 82 is 6.954.

Because, the empiric characteristics (obtained on the basis of the two dispersion estimations), F = 14.779 is smaller than the critical value (6.954) at significance level 0.01, it follows that assessment results give grounds to consider that the Lab work influences significantly on the acquired knowledge and skills of subject “Computer Networks”.

4. CONCLUSIONS

In present paper have been proposed labs for education on the subject “Computer Networks”, as well as statistical analysis of assessment results which carried out two students groups from Southwest University– Blagoevgrad.

The last section of present paper evaluates how much valuable results at learning have been achieved using these labs through dispersion analysis of the examination results of students. The analysis give grounds to consider that the education with combined application of lectures and proposed Labs gave generally the best results.

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PREPARATION OF JINR TO THE DISTRIBUTED DATA PROCESSING OF THE ATLAS EXPERIMENT AT LHC

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Abstract: *The ATLAS experiment on the Large Hadron Collider at CERN will start operation in the end of 2007. The amount of experimental data is expected to be enormous, and will reach 3.5 PB/year. To process such amount of data, the distributed computing*

system based on Grid is being constructed. Joint Institute for Nuclear. Research participates this activity as part of Russian Tier-2 federation. Details of distributed data processing of the ATLAS experimental data are described. Status of preparation work carried out at JINR is presented. Distributed data analysis techniques are explained.

Keywords:*JINR, ATLAS, Grid, high energy physics, distributed analysis*

1 INTRODUCTION

The main goal of the ATLAS experiment[1] is to investigate various aspects of elementary particle properties in high-energy interactions of protons at the Large Hadron Collider (LHC)[2], which is being built at the European Laboratory for Particle Physics (CERN). The experiment is expected to start operation in 200. Data flow from the detector is estimated to reach 3.5 PB/year[3]. The experiment's data processing includes event reconstruction, simulation and physics analysis.

Due to large amount of data to be processed, the ATLAS Computing Model[3] assumes that computing resources are distributed across multiple locations.The tier structure is implemented to combine distributed resources, with distinct roles of the various tiers.

Joint Institute for Nuclear Research (JINR) participates Russian ATLAS Tier-2, in scope of EGEE-RDIG project[4]. According to the Tier-2 role, JINR computing facilities should provide resources for the Monte-Carlo simulation and analysis capacity for physics working groups.

2 THE ATLAS DATA MANAGEMENT

The main requirement on the ATLAS Computing Model is to enable all members of the ATLAS Collaboration an access to the experimental data during the data taking period. The experimental data are stored in a number of data formats. Each data format describes the physics event representation which is successively derived from experimental of simulation data.

RAW data are physics events as output of the ATLAS data acquisition system, and contain primary experimental measurement. Amount of data in RAW format is expected to reach 3.5

PB/year. Reconstruction of RAW data results in Event Summary Data (ESD), which contain detailed physics description of events. ESD data are also stored, and amount of ESD is estimated to be about 1 Pb/year.

Not all information containing in ESD data is necessary for physics analysis, except for some specific tasks. To reduce amount of data to be processed during analysis, the Analysis Object Data (AOD) format is introduced. AOD data are derived from ESD data, and contain information of analysis interest. Amount of AOD data is about 200 TB/year. To allow

fast selection of events of interest, the event metadata are derived from AOD format, and stored separately as TAG data. Amount of TAG is about 2 TB/year. Result of the Monte-Carlo simulation is stored in the same format as experimental data.

The primary RAW experimental data received from the ATLAS detector are processed and stored at Tier-0 at CERN. 10 Tier-1 computing centres are responsible for reconstruction of events and storage of resulting ESD, AOD and TAG data and backup copy of RAW data. AOD data distributed then across about 30 Tier-2. Each Tier-2 is responsible for storage of 1/3 of full set of AOD data and full copy of TAG, and provision of access to the data and computing resources for analysis to any member of the Collaboration.

Given the worldwide distribution of AOD data, the data analysis must be distributed as well. The distributed analysis system should enable physicists to submit jobs from any location with processing to take place at the remote sites. Complete results are expected at the end of job, while partial results are available during processing. The distributed analysis system will typically split the user's job to process data in parallel. For performance reason, the analysis jobs should be submitted to the computing centres, where relevant data are stored. This requires special tools, which use ATLAS software, grid facilities provided by the LHC Computing Grid (LCG) project[5] and dedicated databases, containing information on the data location. Several such tools are being developed by the ATLAS community, and most advanced are PanDA and GANGA[6]. Since GANGA is mostly oriented to function in the LCG environment, it was adopted at JINR as a default tool of data analysis.

3 Participation of JINR in the ATLAS Tier-2 of the Russian Data-Intensive Grid

JINR participates in the Russian Data-Intensive Grid Consortium, which is organised in scope of the EGEE and the LCG projects. In the ATLAS Computing Model RDIG plays role of Tier-2 centre. JINR is expected to provide by the end of 2007 computing resources for the LHC data processing, both for simulation and data analysis. Resources requested are summarised in Tab.1. The existing JINR Computing Centre is being upgraded accordingly at the moment.

Currently, resources of JINR grid segment are represented by of two computing elements (CE) of 42 CPUs in total, and by two storage elements (SE) with combined capacity 40 TB of disk space. Storage elements are operated by DPM and dCache storage systems. Resource broker, monitoring and proxy nodes are available. The grid segment is running gLite-3.0 middleware.

Table1: Resource requirements

CPU	1000 kSi2K
Data storage on disk	200 TB
Data storage on tape	
Network bandwidth	2.5 Gbps

Preparation to the ATLAS distributed data processing is under way. JINR site is integrated into the ATLAS Distributed Data Management System (DDM). Currently, it allows to subscribe to the Monte-Carlo simulation datasets, available at relevant Tier-1 centre, exactly in the same way as it should be done for experimental data, when LHC starts operation.

Intensive work is ongoing among JINR physicists - participants of ATLAS experiment, to train and practise with the GANGA distributed analysis tool. Dedicated job queue has been set up on one of JINR's CE, to allow fast processing of analysis jobs.

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A NEW APPROACH TO EMBEDDED APPLICATIONS BASED ON MICROCONTROLLERS USE USB INTERFACE TO COMMUNICATE WITH PC

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***Abstract**: This paper presented one possibility to development a simple and small component count USB data acquisition system, used Delphi on PC side to communicate. The main core of USB device is PIC18F4550. This device affords an opportunity to switch different sensors and registered their evidences via USB interface on PC. All received data was storages in .txt file, with standard structure. The software support for the data acquisition system device includes the sources of the program on the chip and the Windows application communicated with device as well as the driver. The real device was produced, tested and analyzed. The results are presented in this article.*

***Keywords:** Data acquisition, USB, software.*

1. INTRODUCTION

The RS-232 serial interface is no longer a common port found on a personal computer (PC). This is a problem because many embedded applications use the RS-232 interface to communicate with external systems, such as PCs. A solution is to migrate the application to the Universal Serial Bus (USB) interface. There are many different ways to convert an RS-232 interface to USB, each requiring different levels of expertise. The simplest method is to emulate RS-232 over the USB bus. An advantage of this method is the PC application will see the USB connection as an RS-232 COM connection and thus, require no changes to the existing software. Another advantage is this method utilizes a

Windows® driver included with Microsoft® Windows® 98SE and later versions, making driver development unnecessary. The objectives of this application note are to explain some background materials required for a better understanding of the serial emulation over USB method and to describe how to migrate an existing application to USB. A device using the implementation discussed in this paper shall be referred to as a USB RS-232 emulated device. All references to the USB specification in this document refer to USB specification revision 2.0. Features in version 1.0 of the RS-232 Emulation firmware:

- A relatively small code footprint of 3 Kbytes for the firmware library
- Data memory usage of approximately 50 bytes (excluding the data buffer)
- Maximum throughput speed of about 80 Kbytes
- Data flow control handled entirely by the USB protocol
- Does not require additional drivers; all necessary files, including the .inf files for Microsoft® Windows® XP and Windows® 2000, are included.

2. OVERVIEW

A Windows application sees a physical RS-232 connection as a COM port and communicates with the attached device using the CreateFile, ReadFile, and WriteFile functions. The UART module on the PICmicro® device provides an embedded device with a hardware interface to this RS-232 connection. When switching to USB, the Windows application can see the USB connection as a virtual COM port via services provided by two Windows drivers, usbser.sys and ccport.sys. In-depth details regarding these Windows drivers are outside the scope of this document. A virtual COM port provides Windows applications with the same programming interface; therefore, modification to the existing application PC software is unnecessary. The areas that do require changes are the embedded hardware and firmware. For hardware, a microcontroller with an on-chip full speed USB peripheral is required to implement this solution. The PIC18F4550 family of microcontrollers is used here as an example. References to the device data sheet in this document apply to the “PIC18F4550 Data Sheet” [1]. Firmware modifications to the existing application code are minimal; only small changes are needed to call the new USB UART functions provided as part of the USB firmware framework written in C. Figure 1 shows an overview of the migration path. Migrating to USB using the RS-232 serial emulation method provides the following advantages:

- It has little or no impact on the PC software application
- It requires minimal changes to the existing application firmware
- It shortens the development cycle time
- It eliminates the need to support and maintain a Windows driver which is a very demanding task
- Finally, the method described here utilizes a clear migration path from many existing PICmicro devices to the PIC18F4550 family of microcontrollers, making the upgrade to USB straightforward

Since the USB protocol already handles the details of low-level communication, the concept of baud rate, parity bit and flow control for the RS-232 standard becomes abstract.

Devices in the PIC18F4550 family incorporate a fully featured Universal Serial Bus communications module that is compliant with the USB Specification Revision 2.0. The module supports both low-speed and full-speed communication for all supported data transfer types. This family of devices offers the advantages of all PIC18 microcontrollers – namely, high computational performance at an economical price – with the addition of high endurance, Enhanced Flash program memory. In addition to these features, the PIC18F4550 family introduces design enhancements that make these microcontrollers a logical choice for many high-performances, power sensitive applications. Devices in the PIC18F4550 family are available in 28-pin and 40/44-pin packages.

3. OVERVIEW OF USB

USB device functionality is structured into a layered framework graphically shown in Fig.1. Each level is associated with a functional level within the device.

The highest layer, other than the device, is the configuration. A device may have multiple configurations. For example, a particular device may have multiple power requirements based on Self-Power Only or Bus Power Only modes. For each configuration, there may be multiple interfaces. Each interface could support a particular mode of that configuration.

The PIC18F4550 device family contains a full-speed and low-speed compatible USB Serial Interface Engine (SIE) that allows fast communication between any USB host and the PIC® microcontroller. The SIE can be interfaced directly to the USB, utilizing the internal transceiver, or it can be connected through an external transceiver. An internal 3.3V regulator is also available to power the internal transceiver in 5V applications. Some special hardware features have been included to improve performance. Dual port memory in the device's data memory space (USB RAM) has been supplied to share direct memory access between the microcontroller core and the SIE. Buffer descriptors are also provided, allowing users to freely program endpoint memory usage within the USB RAM space. A Streaming Parallel Port has been provided to support the uninterrupted transfer of large volumes of data, such as isochronous data, to external memory buffers. Fig. 2 presents a general overview of the USB peripheral and its features.

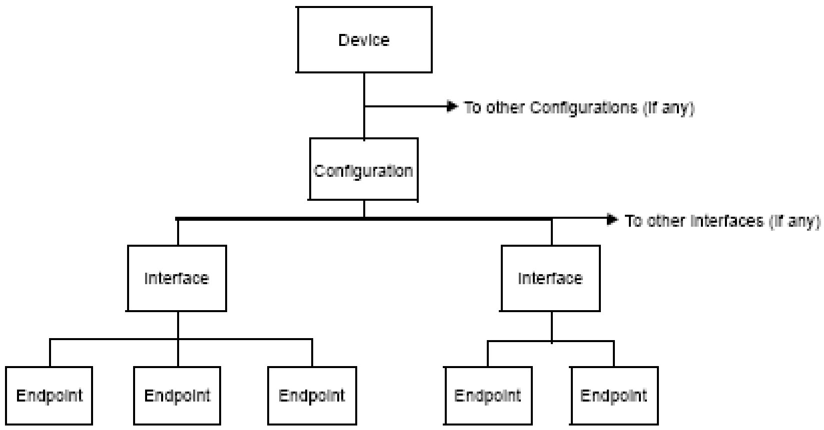

Fig. 1: Usb Layers

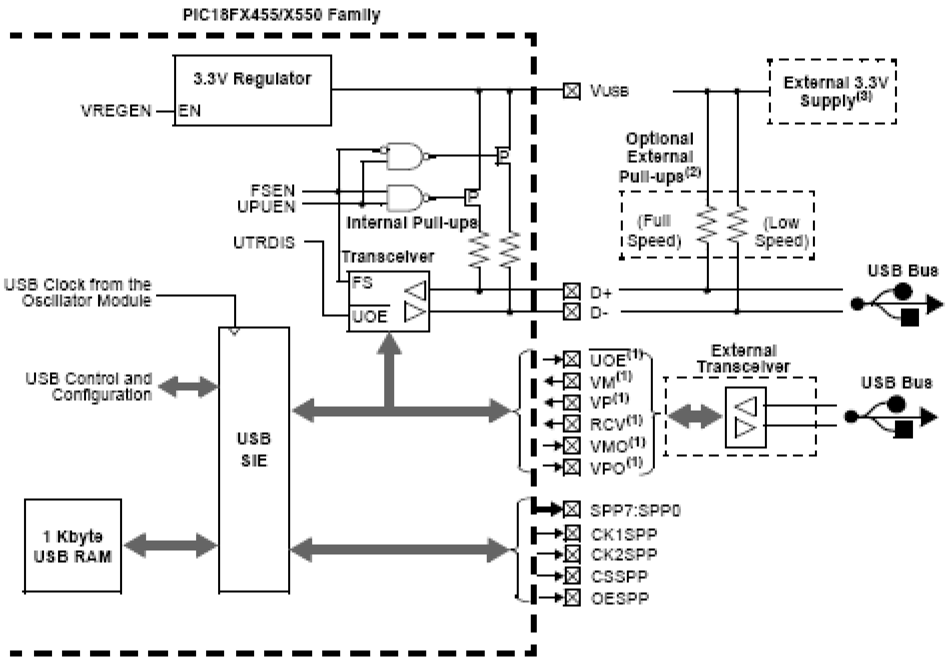

Fig. 2: Usb peripheral and options

The Communication Device Class (CDC) specification defines many communication models, including serial emulation. All references to the CDC specification in this document refer to version 1.1. The Microsoft Windows driver, usbser.sys, conforms to this specification. Therefore, the embedded device must also be designed to conform with this specification in order to utilize this existing Windows driver. In summary, two USB interfaces are required. The first one is the Communication Class interface, using one IN interrupt endpoint. This interface is used for notifying the USB host of the current RS-232 connection status from the USB RS-232 emulated device. The second one is the Data Class interface, using one OUT bulk endpoint and one IN bulk endpoint. This interface is used for transferring raw data bytes that would normally be transferred over the real RS-232 interface. Designers do not have to worry about creating descriptors or writing function handlers for Class-Specific Requests. Descriptors for a USB RS-232 emulated device and all required handlers for Class-Specific

4. SHEMATIC

This USB Data acquisition enable 8 Digital output, 8 Digital input, 8 Analog output. The device no external power required. The below given circuit in fig.3.

Devices in the PIC18F4550 family incorporate a different oscillator and microcontroller clock system than previous PIC18F devices. The addition of the USB module, with its unique requirements for a stable clock source, make it necessary to provide a separate clock source that is compliant with both USB low-speed and full-speed specifications. To accommodate these requirements, this device include a new clock branch to provide a 20 MHz clock for full-speed USB operation.

When connect this USB Data acquisition with PC then windows ask for driver after driver installed(only first time) you will have a new COMx if not connect the COMx not create by Windows.

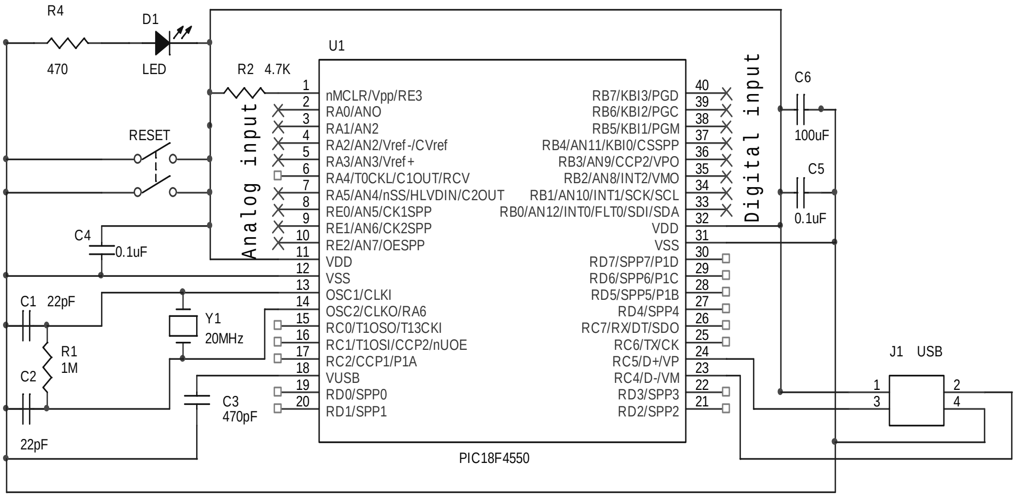

Fig. 3: Schematic of the USB Data acquisition.

Power is available from the Universal Serial Bus. The USB specification defines the bus power requirements. Devices may either be self-powered or bus powered. Self-powered devices draw power from an external source, while bus powered devices use power supplied from the bus. The USB specification limits the power taken from the bus. Each device is ensured 100 mA at approximately 5V (one unit load). Additional power may be requested, up to a maximum of 500 mA. Note that power above one unit load is a request and the host or hub is not obligated to provide the extra current. Thus, a device capable of consuming more than one unit load must be able to maintain a low-power configuration of a one unit load or less, if necessary. The USB specification also defines a Suspend mode. In this situation, current must be limited to 500 μA, averaged over 1 second. A device must enter a Suspend state after 3 ms of inactivity (i.e., no SOF tokens for 3 ms). A device entering Suspend mode must drop current consumption within 10 ms after Suspend. Likewise, when signaling a wake-up, the device must signal a wake-up within 10 ms of drawing current above the Suspend limit.

5. WINDOWS APPLICATION AND COMUNICATION

While the provided USB UART functions greatly simplify the integration of USB into an application, there are still some considerations to keep in mind when developing or modifying application code.

5.1. Code design

The Microchip USB firmware is a cooperative multitasking environment. There should not be any blocking functions in the user code. As USB tasks are polled and serviced in the main program loop, any blocking functions that are dependent on the state of the USB may cause a deadlock. Use a state machine in place of a blocking function.

mUSBUSARTTxRom and mUSBUSARTTxRam expect a data pointer of type rom byte* and byte*, respectively. Type casting may be necessary. 3. while(!mUSBUSARTIsTxTrfReady()); is a blocking function. Do not use it.

putrsUSBUSART, putsUSBUSART, mUSBUSARTTxRom and mUSBUSARTTxRam are not blocking functions. They do not send out data to the USB host immediately, nor wait for the transmission to complete. All they do is set up the necessary Special Function

Registers and state machine for the transfer operation. The routine that actually services the transfer of data to the host is CDCTxService(). It keeps track of the state machine and breaks up long strings of data into multiple USB data packets. It is called once per main program loop in the USBTasks() service routine. Because of this, back-to back function calls will not work; each new call will override the pending transaction.

A correct set of descriptors for the CDC class must be used. Refer to the reference design project for an example.

The endpoint size for the data pipes are defined by CDC_BULK_OUT_EP_SIZE and CDC_BULK_IN_EP_SIZE, located in the header file usbcfg.h. Since these endpoints are of type bulk, the maximum endpoint size described must either be 8, 16, 32 or 64 bytes.

The type byte is defined as an unsigned char in the header file typedefs.h.

Always check whether the firmware driver is ready to send more data out to the USB host by calling mUSBUSARTIsTxTrfReady().

5.2. Setting up the code project

1. Insert #include "system\usb\usb.h" in each file that uses the CDC functions.

2. USB_USE_CDC should be defined in the file usbcfg.h when using the CDC functions.

3. The source and header files, cdc.c and cdc.h, should be added to the project source files. They can be found in the directory “system\usb\class\cdc\”.

5.3. USB Vendor ID (VID) and Product ID (PID)

The VID and PID are important because they are used by the Windows operating system to differentiate USB devices and to determine which device driver is to be used. The VID is a 16-bit number assigned by the USB Implementers Forum (USB-IF). It must be obtained by each manufacturer that wants to market and sell USB products. The VID can be purchased directly from USB-IF. More detailed information can be found at: http://www.usb.org/developers/vendor. Each VID comes with 65,536 different PIDs which is also a 16-bit number. In the Microchip USB firmware framework, the VID and PID are located in the file usbdsc.c. Both values can be modified to match different product VID and PID numbers.

5.4. Drivers for Microsoft Windows® 2000 and Windows® XP

Microsoft Windows does not have a standard .inf file for the CDC driver. The drivers are, however, part of the Windows installation. The only thing necessary to do is to provide an .inf file when a CDC device is first connected to a Windows system.

Example .inf files are provided with the CDC RS-232 Emulation Reference Project and are located in the source code directory <Install>\fw\CDC\inf. Before using them, they must be modified to reflect the application’s specific VID and PID. This is in addition to any changes to usbdsc.c that have already been made and must match those values. The VID and PID are located in the string “USB\VID_xxxx&PIDyyyy”, where “xxxx” is the hexadecimal VID and “yyyy” is the hexadecimal PID. The string is generally part of one of the lines under the heading “[DeviceList]”. If desired, users may also modify the variable definitions under the heading “[Strings]”. This changes the device identification text seen by the user in the device manager.

6. SUMMARY

The RS-232 serial emulation provides an easy migration path for applications moving from UART to USB. On the PC side, it requires minimal software modification. On the embedded device side, the PIC18F4550 family of microcontrollers provides a simple hardware upgrade path from the PIC16C745/765 and PIC18FXX2 families of devices. Library firmware with user-friendly APIs are also included for convenience. Tutorial exercises are available as part of the CDC RS-232 Emulation Reference

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PROGRAM SYSTEM FOR INVESTIGATION OF HEAT PHYSICS APPLICATIONS

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Abstract: *Software application for heat analysis of technical devices and systems is developed and experimented. It is based on detailed algorithm for numerical solving of transient three dimensional heat transfer equation with various types of boundary condition. The program application 'Eart_Hea't is developed for using in solving different research problems (energy balance in buildings, heat transfer in earth layers, heat accumulators, solar applications, greenhouses e.g.).*

Keywords: *Heat transfer modeling, finite difference method, numerical analysis.*

1. INTRODUCTION

Maintaining a comfortable temperature inside buildings or other objects, like greenhouses and pools, require a significant amount of energy. Considering that 46% of sun's energy is absorbed by the earth, a good option is to use this abundant energy to heat and cool buildings and other objects. In contrast to many other sources of heating and cooling energy which need to be transported over long distances, Earth Energy is available on-site, and in massive quantities.

Because the ground transports heat slowly and has a high heat storage capacity, its temperature changes slowly—on the order of months or even years, depending on the depth of the measurement. As a consequence of this low *thermal conductivity*, the soil can transfer some heat from the cooling season to the heating season; heat absorbed by the earth during the summer effectively gets used in the winter.

This warm earth and groundwater below the surface provides a free renewable source of energy that can easily provide enough energy year-round to heat and cool an average suburban residential home, for example.

The main problem for assessing the heat transfer equipment for Earth Energy utilization is very complicated heat transfer mechanism. In charge and discharge phase of the process there is continuous change of heat flux. This mean that the energy extracted and accumulated in soil will vary. Variation of extracted or accumulated heat can be calculated only by using tree-dimensional transient mathematical model with distributed heat sources (heating or cooling). Computer programming for such a calculation problems require serious algorithmic organization for achieving the needed universality for different tasks.

2. THEORETICAL MODEL

Mathematical model of processes in ground layers can be derived on the base of theoretical treating of heat conductivity, heat transfer to the working fluid and heat losses to the ambient. The main mathematical equation in this problem is the heat conductivity equation, written in orthogonal coordinate system:

(1) $$\left(\frac{\partial^2 T}{\partial x^2}+\frac{\partial^2 T}{\partial y^2}+\frac{\partial^2 T}{\partial z^2}\right)+\frac{q_V}{\lambda}=\frac{1}{a}\frac{\partial T}{\partial t}$$

where **T** is the temperature in the walls**,** ***x, y*** and ***z*** – space variables, ***λ*** – heat conductivity coefficient [W/m K], q_V - source member [W/m^3] and $a=\lambda/\rho\cdot c_p$ is thermal diffusivity.

This equation is first order in time and second order in space, so it require one boundary condition in time (called an initial condition) and surrounding boundary conditions in space - 6 boundary conditions for 3D problem. In dependency of boundary condition there can be defined different tasks describing real heat processes. For example, there can be solved the one-dimensional problem for temperature distribution in massive building walls, which are the base conception of so-called passive solar systems. Two and three-dimensional models can be used for accumulating and extracting the heat from massive earth volume.

The mathematical problem is defined on a rectangular domain [0; L_x]x[0; L_y]x[0; L_z]x[0; T] and is assumed that L_x, L_y, L_z and T are chosen properly for practical purposes.

3. BOUNDARY CONDITIONS

In general, boundary conditions for objects considered in this work, can be written in next form:

(2) $$\beta\frac{\partial T}{\partial n}+\gamma T+\sigma=0$$

where ***β,γ*** and ***σ*** are coefficients, which present heat conductivity, heat convective coefficient to the free space, and heat sources (for example solar radiation); ***n*** - normal direction to the boundary surface.

The main problem in supporting the boundary condition data for the model is connected with the weather parameters (solar radiation, ambient temperature). Base concept for weather data is the availability of temperature bins for daytime and nighttime's hours for the selected location. Additionally, bin data for the coldest and hottest months (corresponding to design heating and cooling conditions) are required for the ground loop calculation. Such a heavy user-data requirement would render the model impractical. Alternatively, storing the data within the model would translate into an excessively large file if even a moderate number of locations around the world were to be included.

To circumvent this problem, an hourly weather data generator is included in the model. This generator is based on empirical correlations between statistically derived daily maximal and minimal ambient temperature and hourly weather data, as defined in ASHRAE (1997). Usage of a generator does not restrict the generality of the method. If appropriate bin data are available, they could be used in the model without any change to the other algorithms.

4. FINITE DIFFERENCE APPROXIMATION

Because of the importance of the diffusion/heat equation to a wide variety of fields, there are many analytical solutions of that equation for a wide variety of initial and boundary conditions. However, one very often runs into a problem whose particular conditions have no analytical solution, or where the analytical solution is even more difficult to implement than a suitably accurate numerical solution. The finite difference method begins with the discretization of space and time such that there are an integer number of points in space and an integer number of times at which is calculated the field variable(s), in this case just the temperature. The resulting approximation is shown schematically in figure 1. For simplicity here is described only one dimension mesh with intervals of size $\Delta x = x_{i+1} - x_i$.

The finite-difference form of differential equation (1) is derived by integration over control volume surrounded the typical node ***i, n*** in solution grid (Fig.1). The indexes ***I,j,k*** and ***n*** refer to the thickness (***x,y,z***) and the time (τ). An implicit time approximation, which is stable for forward integration in time, is developed for transient differential equations. If the time interval is named $\nabla\tau = (\tau_n, \tau_n + \nabla\tau_{n+1})$, the time derivative can be written using forward Euler formula for discretization:

(3) $$\frac{\partial T}{\partial \tau} = \frac{T^{n+1} - T^n}{2\nabla\tau}$$

For the space derivative is applied a second order nonlinearly implicit Crank-Nicholson method, which is solved by iteration.

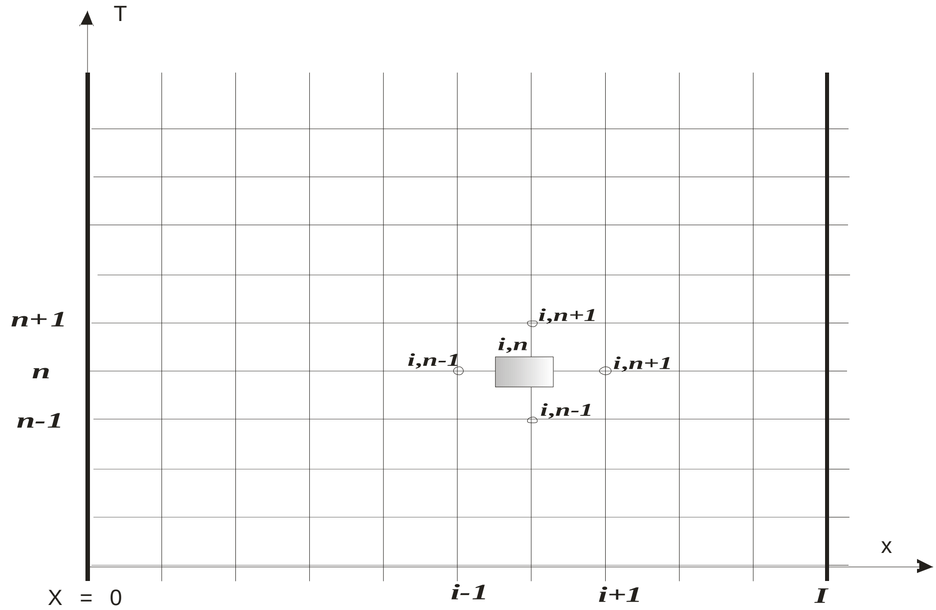

Fig. 1 The mesh in time and space

(4) $$\frac{\partial^2 T}{\partial x^2} = \sigma \frac{T^{n+1}_{i+1,j,k} - 2T^{n+1}_{i,j,k} + T^{n+1}_{i-1,j,k}}{\Delta x^2} + (1-\sigma)\frac{T^{n}_{i+1,j,k} - 2T^{n}_{i,j,k} + T^{n}_{i-1,j,k}}{\Delta x^2}$$

where σ is a weight coefficient.

This is a system of (***I***-2) algebraic equations with ***I*** (***i*** = 1,2....***I***) unknown node temperatures (with upper index ***n+1***). Temperatures with upper index ***n*** are considered as known (received from calculation, made in former time step or from initial conditions in the first time step). Boundary conditions (2) must be added to complete the system.

Such equations can be written for Y and Z directions (j and k indexes respectively).

Equations (4) can be solved by standard algebraic methods. For solving the three dimensional problem is needed to add equations like (4) for other directions Y and Z. This is made by using so called fractional steps method. It is equivalent to separate solving the one-dimensional task for each time step.

5. PROGRAM EARTH_HEA T

It is developed a software system EARTH_HEAT for modeling and simulation calculations of thermal processes of heat extracting and accumulation in earth. The algorithm and the organization of the interface of the program provides a universal approach for adding new type of devices, including an opportunity for variation of parameters for the elements in the system.

The various algorithms are used to calculate, on a month-by-month basis, the energy transfer in systems, utilizatied heat from earth. A flowchart of the algorithm used in EARTH_HEAT system is shown in Figure 2.

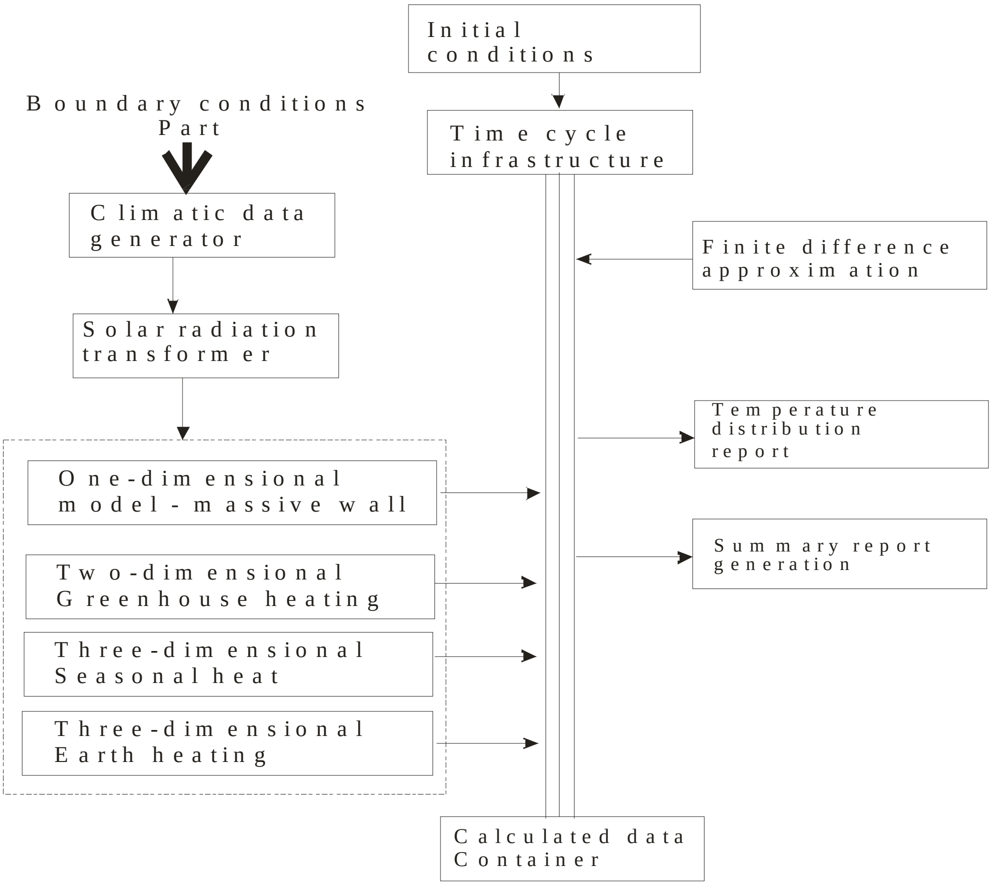

Fig. 2. Scheme of the program algorithm

The heat transfer in Earth_Heat systems is relatively complex because of great number of factors influencing the boundary condition data. It is dependent upon the solar radiation, temperature and wind speed surrounding the system. Most heating analyses tools use an hourly time step to follow the changing solar and weather conditions.

The organization of the programming system contains three basic functional parts (Fig. 1). As a basic infrastructural part of the system it appears the module of controlling simulation cycles in time and providing necessary climatic parameters in relation to discretion of the processes in time. In this time it is included the deliverance of data about the solar radiation and the air temperature for different geographical regions. In this functional part it is included special solar radiation' processor, which recalculate solar radiation for given slope and orientation of the received surface, and optical parameters for solar radiation penetration through the transparent covers.

The control of the system is carried out by a main controlling form (panel) which delivery the boundary condition data (fig.2). Through this panel it can be chosen the source of climatic data about simulation calculations and the type of the installations for utilizing earth energy. In the performed variant of the program there is a possibility to be carried out thermo-technical analyses of devices (heating of building, heat accumulator in soil greenhouse heating, passive solar system), and also additional analyses about the parameters of the temperature distribution in earth.

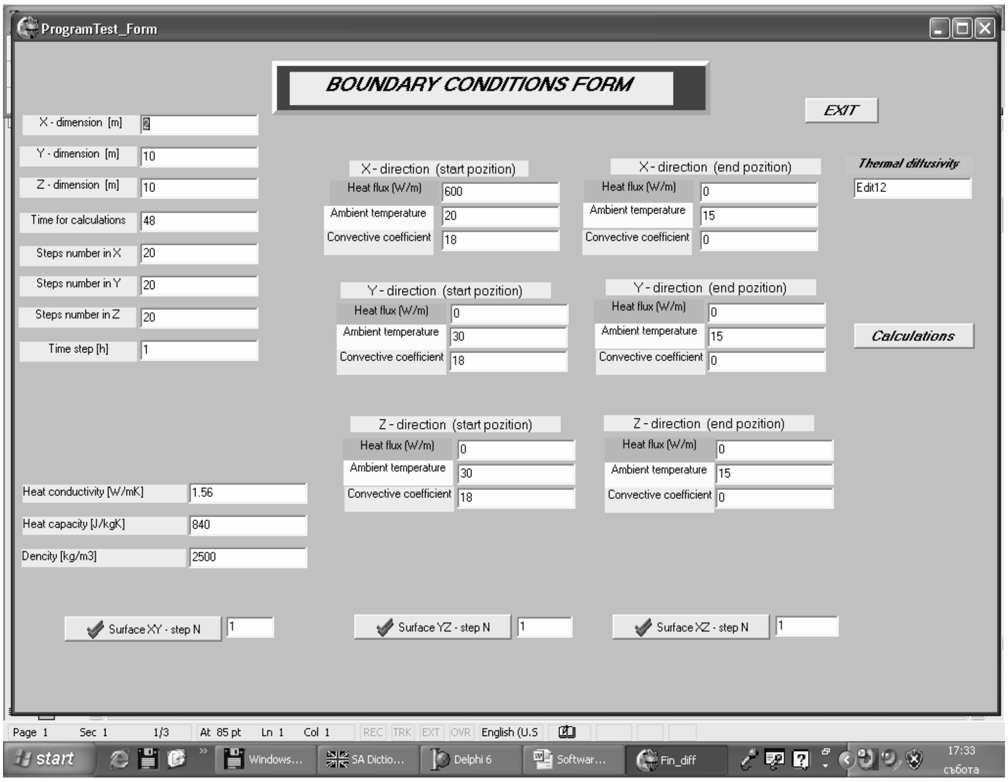

Fig. 2 Program form "Boundary conditions"

6. CONCLUSIONS

A software product has been performed for simulation analyses of devices for thermal transformation of energy from earth. It has modular structure for completing different installation schemes and it gives the opportunity to generate long-term assessments of thermal efficiency of devices. Program system can be successfully used not only for projecting purposes, but also for solving exploring tasks.

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PROBABILITY-INFORMATIONAL MODEL OF MEASUREMENT

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***Abstract**: The measurement is an information process and the measuring instrument is an information tool. During the measurement the unknown variable is quantatively evaluated by using a measurement unit. The International System of Units (SI) was adopted covering basic, supplementary and derived units. Each measurement is performed with some error and each measuring instrument possesses certain error. There are several sources and several categories of error. When evaluating an instrument its static and dynamic characteristics and errors should be carefully examined. For high-accuracy measurements statistical methods of measurement data processing are recommended.*

1. INTRODUCTION

For the classical conversional scheme of measurement means, shown in figure 1

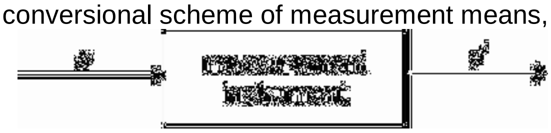

Fig. 1

where x is the input quantity, y' - the output quantity, the equation $y' = \varphi(x)$ presents a statical characteristic. From here it can be obtain the reverse statical characteristic $x = \alpha(y')$, named also scaling.

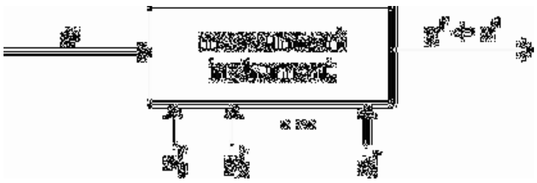

Fig. 2

In fact the same conversional scheme has a kind, shown in figure 2, when $y = \alpha(y' + z') \neq x$, i.e. the error z can be presented as $z = y - x$. The generalized set model of measurement by measurement means is shown in figure 3.

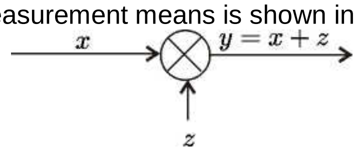

Fig. 3

In this respect the measurement of one realization of a random quantity with known distribution graphically can be pictured like figure 4.

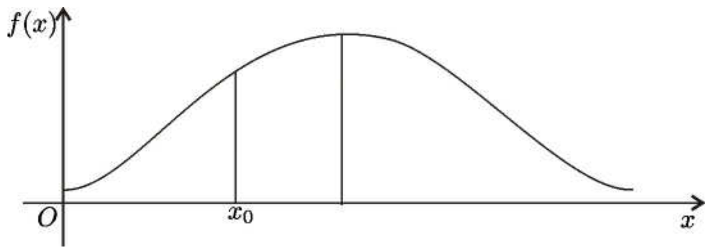

Fig. 4

But the measurement of a constant, determining (not random) quantity will be viewed as figure 5.

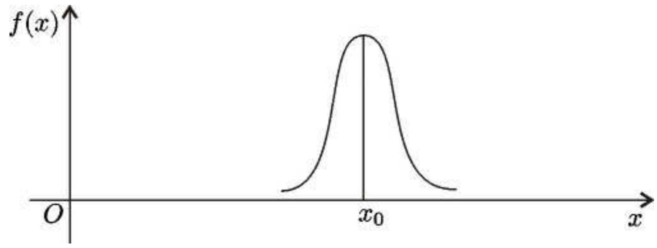

Fig. 5

Here $f(x) = \delta(x - x_0)$. At $x = x_0$, $\delta(x - x_0) = \delta(0) = 1$.

The second (classical) case can be reduce to the first with the unconditional probabilities

$$\begin{array}{ccccc} Y & = & X & + & Z \\ \downarrow & & \downarrow & & \downarrow \\ f(y) & & f(x) & & f(z) \end{array}$$

This is the same mathematical model of the measurement of realization of the random quantity X .

The peculiarity consist of the fact that the result Y is connected with X by probability relationship when Z and X can be dependent random quantity also.

Hence, it can be done the next conclusion: under the *measurement of a random quantity* it will be understand the measurement of the realization of the random quantity.

2. ESTIMATE OF PROBABILITY CHARACTERISTICS

The probability characteristics presenting non random numbers or functions are estimated by an ensemble from infinite number of realization or from one realization with infinite continuance if the random process is stationary and ergodic.

In practice the number of realizations used for experimental investigations or the continuance of one realization of the stationary ergodic random process (the time for observation) always are limited. That is way in fact every statistical characteristic, obtained by a hardware differs from probability (theoretical) characteristic, presented an object of measurement.

The obtained statistical characteristic can be assumed as true one and to be named estimate of the measured probability characteristic.

3. STATISTICAL ERROR ANALYSES

The In the common case, the classic approach distinguishes two components of the measurement error Δ; say, a systematic Δ_S and a random $\overset{o}{\Delta}$ error:

$$\Delta = \Delta_s + \overset{o}{\Delta},$$

where

Δ_s - Systematic error,

$\overset{o}{\Delta}$ - Random error.

The systematic error can be identified and eliminated by correction.

The random component of the error can be evaluated by a suitable mathematical treatment of the measurement results.

As mentioned the aim of statistical error analysis is to evaluate the random measurement error, thus to determine the most probable magnitude of the measurand as well as the most probable random error.

The sciences describing and evaluating both random events and random quantities are called Probability theory and Mathematical statistics.

In the measurement practice the information about a measurand is referred to as sample: $x_1, x_2, ..., x_n$.

Every random quantity is best characterized by its own distribution law.

The function $P(x)$, which gives the relationship between the values x_i of the variable random quantity X and the probabilities for their appearance p_i is called distribution law of the quantity.

The continuous variable has infinitely large uncountable sets of values even for a comparatively small interval.

Therefore, the probability distribution laws are discrete or continuous.

The random error inherent to the measurement process is a continuous quantity.

As with discrete probability distributions, the sum of the probability values must equal 1. Because there are an infinite number of values of the variables, however, the probability of each value of the variable must be 0. If the probability values for the variable values were greater than 0, then the sum would be infinitely large.

In this case we can report only these values which can be assumed with one or another probability in a given interval "from" and "to". This interval can be as large, as well as narrow. If this interval is enough narrow: from *x* to *x* + *dx*, and if there is not dominating and strongly determining factors into whole complex of conditions, necessary for the execution of a given measurement procedure, then the distribution law is referred to as normal (or Gaussian) law and the probability is given by the equation:

$$dP(x) = W(x).dx = \frac{1}{\sigma\sqrt{2\pi}} e^{-\frac{1}{2}\frac{(x-\mu)^2}{\sigma^2}}.dx,$$

where

dx - is a small quantity, defining the interval width,

π, e - are well known mathematical constants,

σ - is square mean deviation, characterizing the dispersion degree of the values x_i of the random continuous quantity X around the mean value,

μ - named mathematical expectation,

$W(x)$ - probability density.

The probability distribution law is the best description of the random variable, but is not convenient for engineering calculations. In the Probability theory is proved that a distribution law can be defined by numerical characteristics. These numerical characteristics can be used for estimation of the random quantities (i.e. for quantitative estimation of the random errors).

These numerical characteristics are:

- Mathematical expectation - first (initial) moment

$$M(x) = \int_{-\infty}^{+\infty} x_i.W(x).dx,$$

where

x_i is the random quantity,

$W(x)$ - is the probability density function.

The units of *M(x)* are the same as are the units of *x*.

- Dispersion - second (central) moment

$$D(x) = \int_{-\infty}^{+\infty} [x_i - M(x)]^2 .W(x).dx,$$

where

x_i is the random quantity,

$W(x)$ - is the probability density function.

The dispersion has a dimension "squared" with respect to *x* and therefore is not convenient parameter. Very often the quantity

$$\sigma(x) = \sqrt{D(x)}$$

is used. This quantity is referred to as square mean value. The units of σ are the same as are the units of *x*.

Such parameters are determined for $n \to \infty$ and are referred to as parameters of the general population, which is not practical from engineering point of view.

Usually, in the measurement practice we work with limited samples and the characteristics we obtain in this case have to be considered as estimators of the corresponding parameters, i.e. instead of *M(x)* we work with the estimator - $\overline{M}$, instead of $D(x)$ we work with the estimator $\overline{D}$, also instead of σ we work with the estimator *S*.

The nearness of the estimators to the corresponding characteristics is called reliability of the estimators. The reliability of a given estimator very often is determined by using of the parametrical criterion - Student's - *t* - criterion (*t*- distribution), under the pen name "Student" of the English mathematician W. S. Gosset.

According Gosset

$$t = \frac{\overline{X} - \mu}{\sigma / \sqrt{n}},$$

when n is the number of measurements respectively

$$t = \frac{\overline{X} - X_r}{S / \sqrt{n}} = \pm t(p,n),$$

or

$$X_r = \overline{X} \pm t(p,n)\frac{S}{\sqrt{n}},$$

where

t(p,k) are given in tables for several values of confidence level $\alpha = p - 1$ and degrees of freedom $k = n - 1$.

The choice of confidence level should be made by the user of the measurement results in accordance with technical or economic factors of which he may be aware.

As a measure of location, the estimation of μ is generally obtained by calculating the arithmetic mean $\overline{X}$ of a set of *n* measurement results $x_1, x_2, ..., x_n$, i.e.

$$\overline{X} = \frac{1}{n}\sum_{i=1}^{n} x_i .$$

The symbol S denotes a sample standard deviation. Thus,

$$S = \sqrt{\frac{\sum_{i=1}^{n}(x_i - \overline{X})^2}{n-1}}$$

is an estimate of the general standard deviation σ.

Hence, the notion "interval estimation" is given as:

$$\{X_r\} = \overline{X} \pm t(p,n)\frac{S}{\sqrt{n}},$$

where $\overline{X}$ can be considered as a "point estimation".

F. E. Grubs in 1950 suggested a method for finding out of gross errors in the measurement results. Such method is applicable under the analysis of samples, for which a hypothesis for normal distribution is proposed.

The extremal values are subjected to a test as follows:

$$\frac{|X_{extr} - \overline{X}|}{S} > \gamma(p,n),$$

where

$\gamma(p,n)$ are given in tables for several values of confidence level and numbers of measurements.

If the test for existing of gross errors is confirmed, then the corresponding result X_{extr} should be eliminated of further processing.

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A SURVEY ON EFFECTIVENESS OF THE PARAMETRICAL ALGORITHM OF PATTERN RECOGNITION

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Abstract: *The paper deals with a parametric algorithm A(m,α, α1,…,δ1, δ2) for calculating evaluations of a set of standard tables (matrices). The tables are generated by normal distribution of random quantities, which are initiated by a measure for an effectiveness of the algorithm A(m,α, α1,…,δ1, δ2). We will prove that (by small enough dispersion) the algorithms of the class under investigation have an extremely high effectiveness on a subclass of the considered class of standard tables.*

Keywords: *pattern recognition, parametrical algorithm.*

1. FORMULATION OF THE PROBLEM

Let us consider metric spaces $\sigma_1, \sigma_2, \ldots, \sigma_n$ with corresponding metrics $p_1, p_2, \ldots, p_n$ ($p_i(\alpha, \beta) = |\alpha - \beta|$). We say that the vector $\mathbf{S} = (s_1, s_2, \ldots, s_n)$ is *an admissible object* (table or matrix row), if $s_i \in \sigma_i, \quad i = 1,2,\ldots, n$.

Let us denote with D the set of admissible objects. We suppose that there exist a separation of *l* non-intersecting classes of subsets (patterns) $K_1, K_2, \ldots, K_l$ for the set *D*.

Let a finite set $T_1 = \{S_1, S_2, \ldots, S_{m_l}\}$, be defined, such that

$$T_1 \subseteq D, \quad T_1 \cap K_i \equiv \{S_{m_{i-1}+1}, S_{m_{i-1}+2}, \ldots, S_{m_i}\} \neq \emptyset, \quad i = 1,2,\ldots, l, \quad m_0 = 0.$$

For given admissible vector ***S*** we want to find the class where ***S*** belongs. The solution must be based on the information from the set of elements of the table T_1.

1.1. Algorithm and basic notions

We propose the following algorithm [1, 2]:

1. Let $\tilde{\Omega}$ be the set of all subsets of the set $\{1,2,\ldots,n\}$. We separate certain set Ω $(\Omega \subseteq \tilde{\Omega})$ (N is a cardinality of Ω): and call it system of support sets for this algorithm. With each support set we associate *n*-dimensional binary vector $\tilde{\omega}$, whose set of unit coordinates, coincides with given support set. We say that the binary vector $\tilde{\omega}$ generates corresponding support set $M_{\tilde{\omega}}$ in $\tilde{\Omega}$.

2. Let $\mathbf{S} = (s_1, s_2, \ldots, s_n)$ and $\mathbf{S}' = (s'_1, s'_2, \ldots, s'_n)$ are possible rows (objects), and $M_{\tilde{\omega}} = \{i_1, i_2, \ldots, i_k\}$ is a support set. The collections $\tilde{\omega}S = (s_{i_1}, s_{i_2}, \ldots, s_{i_k})$, $\tilde{\omega}S' = (s'_{i_1}, s'_{i_2}, \ldots, s'_{i_k})$ are parts of corresponding rows.We define a proximity function $r(\tilde{\omega}S, \tilde{\omega}S')$ for parts of these rows. Let $\alpha_i \geq 0 \quad (i = 1,2,\ldots n); \quad \alpha \geq 0; \quad \alpha \leq m-1$, then

$$r(\tilde{\omega}S, \tilde{\omega}S') = \begin{cases} 1, & \text{if} \quad d \leq \alpha \\ 0, & if \quad d > \alpha \end{cases},$$

where *d* is the number of the non-satisfying inequities $p_i(s_i, s'_{q_i}) \leq \gamma_i, \quad i = i_1, i_2, \ldots, i_k$, p_i is the metric in space σ_i.

3. For each pair of rows (S, S_q), $q = 1,2,\ldots,m_l$, and each support set $M_{\tilde{\omega}} \in \Omega$ is a determined function, called estimator for this pair of rows, based on fixed support set:

$$\tilde{\omega}\Gamma(S, S_q) = f_q^{\tilde{\omega}}(r(\tilde{\omega}S, \tilde{\omega}S_q)), \quad q = 1,2,\ldots,m_l, \quad M_{\tilde{\omega}} \in \Omega.$$

4. For each class $K_j, \quad j = 1,2,\ldots,l$, and each support set $M_{\tilde{\omega}} \in \Omega$ is defined a quantity, called estimator on the same support set:

$$\Gamma_j^s(\tilde{\omega}) = \varphi_j^{\tilde{\omega}}(\tilde{\omega}\Gamma(S, S_{m_{j-1}+1}), \ldots, \tilde{\omega}\Gamma(S, S_{m_j})).$$

5. We specify the ***class estimator*** for system of supports sets as a function of class estimators of each support set:

$$\Gamma_j^s = \psi_j(\Gamma_j^s(\tilde{\omega}_1), \Gamma_j^s(\tilde{\omega}_2), \ldots, \Gamma_j^s(\tilde{\omega}_N)), \quad j = 1,2,\ldots,l, \quad M_{\tilde{\omega}_i} \in \Omega.$$

6. The decision rule is reduced to the calculation of estimators of the function $F(\Gamma_1^s, \Gamma_2^s, \ldots, \Gamma_l^s)$, with values in $\{1,2,\ldots,l\}$. If $F(\Gamma_1^s, \Gamma_2^s, \ldots, \Gamma_l^s) = j, \quad j \neq 0$, then we conclude that, the algorithm classify the row **S** in the class K_j. If $F(\Gamma_1^s, \Gamma_2^s, \ldots, \Gamma_l^s) = 0$, then the algorithm does not recognize **S,** e.g.

$$F\left(\Gamma_1^s,\Gamma_2^s,\ldots\Gamma_l^s\right)=\begin{cases} j,\ \text{if} \begin{cases} 1.^\circ\ \Gamma_j^s-\Gamma_i^s\geq\delta_1, & 1\leq i,j\leq l, i\neq j \\ 2.^\circ\ \dfrac{\Gamma_j^s}{\sum_{i=1}^{l}\Gamma_j^s}\geq\delta_2 \end{cases} \\ 0,\ \text{if this once of the conditions } 1^\circ \text{and } 2^\circ \\ \text{is not realized.} \end{cases}$$

The choice of the set Ω and the functions, defining the algorithm are made in such a way, that:

1. To provide the effectiveness of procedure for calculating estimators $\Gamma_1^s,\Gamma_2^s,\ldots,\Gamma_l^s$ and designating a decision for belongings the object to one or an other class;
2. The corresponding algorithm must be optimal in sufficiently wide class of algorithms.

We consider the finite set $T_2=\{S_1',S_2',\ldots,S_{t_l}'\}$, such that:

$$T_2\subseteq D,\quad T_2\cap K_i\equiv\{S'_{t_{i-1}+1},S'_{t_{i-1}+2},\ldots,S'_{t_i}\}\neq\varnothing,\quad i=1,2,\ldots,l,\quad t_0=0.$$

Let K_{ij} be the set of objects from $K_i'=T_2\cap K_i$, which are classified by the algorithm ***A*** in class K_j, $i,j=1,2,\ldots,l$. Let also K_{i0} be the set of objects from class K_i', $i=1,2,\ldots,l$, which are not recognized from algorithm ***A***.

The value $Q(A)=\dfrac{1}{t_l}\sum_{i=1}^{l}K_{ii}$ is the effectiveness of the algorithm ***A***.

Let a set $\{A\}$ of algorithms for solving recognition problem be given. The algorithm ***A**** is called ***extremal*** if $Q(A^*)=\max_{A\in\{A\}}Q(A)$.

The quality functional gives generalized recognition error in recognizing the control objects (rows):

$$\Phi(A)=\frac{1}{t_l}\left(\sum_{i,j=1}^{l}\alpha_{ij}\left|K_{ij}\right|+\sum_{i=1}^{l}\beta_i\left|K_{i0}\right|\right),$$

where α_{ij},β_i are non-negative penalty constants for incorrect recognition or rejection for K_i', $i=1,2,\ldots,l$. In that case the extremal algorithm ***A**** is specified by the condition: $Q(A^*)=\min_{A\in\{A\}}\Phi(A)$.

The set $\{A\}$ is defined as a one- or multi-parametric family of algorithms in order to ensure efficiency of the implementation of the procedure of finding an extremal algorithm amongst all algorithms of consideration. The set of the parameter values is strictly defined by the set Ω, the functions $f_q^{\tilde{\omega}},\varphi_j^{\tilde{\omega}},\psi_j,F$, $q=1,2,\ldots,m_l$, $j=1,2,\ldots,l$, $M_{\tilde{\omega}}\in\Omega$, and hence by the algorithm. In this case the class of algorithms is one-to-one correspondence with domain defined by multi-parametric space. The problem is to find the absolute extremum of the multi-parametric function over the corresponding domain of multi parametric space.

1.2. Standard tables

The quality of the class of algorithms ***A***(m,α, α1,…,δ1,δ2) is evaluated in regard to the set of pair of tables for training and control. This can be described as:

Let $\eta = (\eta_1, \eta_2, ..., \eta_{2m}, \eta_1', \eta'_2, ..., \eta_{2t}')$ be a system of *2(m+t)* independent *n* -dimensional random vectors:

$$\eta_i = (\eta_{i_1}, \eta_{i_2}, ..., \eta_{i\,n}) \quad i = 1, ..., 2m$$

$$\eta'_j = (\eta'_{j_1}, \eta'_{j_2}, ..., \eta'_{j\,n}) \quad j = 1, ..., 2t$$

whose components are independent and normally distributed corresponding to parameters:

$$M\eta_{l_i} = \tilde{a}_{l_i},\ D\eta_{l_i} = \tilde{\sigma}^2_{l_i},\ l = 1,2,...,\ 2m$$

$$M\eta'_{j_i} = \tilde{b}_{j_i},\ D\eta'_{j_i} = \tilde{\sigma}'^2_{l_i},\ j = 1,2,...,\ 2t \qquad (1)$$

where

$$\tilde{a}_{li} = \begin{cases} a_{1i}, l \le m \\ a_{2i}, l > m \end{cases} \qquad \tilde{\sigma}_{li} = \begin{cases} \sigma_{1i}, l \le m \\ \sigma_{2i}, l > m \end{cases}$$

$$\tilde{b}_{ji} = \begin{cases} b_{1i}, j \le t \\ b_{2i}, j > t \end{cases} \qquad \sigma'_{ji} = \begin{cases} \sigma'_{1i}, j \le m \\ \sigma'_{2j}, j > m \end{cases} \text{, for } i = 1,2,\dots,n.$$

Let us denote the set of random quantities η as Δ. The pair of the tables (T_1, T_2) consists of a table T_1 for training, and a table T_2 for control. This pair (T_1, T_2) realizes the random value $\eta \in \Delta$.

Let $(\beta_1, \beta_2, ..., \beta_{2m}, \beta_1', \beta'_2, ..., \beta_{2s}')$ be an arbitrary realization of a random value $\eta = (\eta_1, \eta_2, ..., \eta_{2m}, \eta_1', \eta'_2, ..., \eta_{2s}')$. Then the corresponding pair of tables (T_1, T_2) is such that T_1 belongs to K^1_r and contains the rows

$$\beta_{m_{j-1}+1}, \beta_{m_{j-1}+2}, ..., \beta_{m_{j-1}+m}, \quad m_{j-1} = (j-1)m, \quad j = 1,2,$$

the table T_2 belongs to K^2_r and contains the rows

$$\eta'_{t_{j-1}+1}, \eta'_{t_{j-1}+2}, ..., \eta'_{t_{j-1}+t}, \quad t_{j-1} = (j-1)t, \quad j = 1,2.$$

In other words – the rows of the class K^i_l, connected to table T_i formed by independent realizations of an $n-$ dimensional random vector $\eta^{(i)}_l$ distributed by the law of normally distribution of random quantities with parameters $\eta^{(i)}_l$, $i = 1,2$:

$$M\eta^i_l = \begin{cases} (a_{l_1}, a_{l_2}, ..., a_{ln}), i = 1 \\ (b_{l_1}, b_{l_2}, ..., b_{ln}), i = 2 \end{cases} \quad D\eta'^i_l = \begin{cases} (\sigma^2{}_{l1}, \sigma^2{}_{l2}, ..., \sigma^2{}_{ln}), i = 1 \\ (\sigma'^2{}_{l1}, \sigma'^2{}_{l2}, ..., \sigma'^2{}_{ln}), i = 2 \end{cases} \quad l = 1,2 \qquad (2)$$

Such a set of pairs of tables for training and control, we call a class of normally distributed tables (for short – normal tables). We assume that the class of normal tables is generated by the set Δ of random variables.

2. A MEASURE FOR THE EFFECTIVENESS OF THE ALGORITHMS A(M,α, $\alpha_1,\ldots,\delta_1,\delta_2$).

Let ***A*** be an arbitrary algorithm of the class of algorithms ***A***(m,α, $\alpha_1,\ldots,\delta_1,\delta_2$) and the pair (T_1, T_2) be an arbitrary realization of the random variable $\eta \in \Delta$. The rows of table T_2 are sequentially classified in accordance to the algorithms ***A***. The value of the functional $Q(A)$ will be fixed, to be equal to the number of correctly recognized control rows. Let A^* be the extreme algorithm from the class ***A***(m,α, $\alpha_1,\ldots,\delta_1,\delta_2$), i.e.

$$Q(A^*) = \max_{A \in A(m,\alpha,\alpha_1\ldots\alpha_n,\delta_1,\delta_2)} Q(A).$$

The value of the functional of quality for the extreme algorithm is a random variable, a function depending on a random variable η. As measure for effectiveness of the algorithm ***A***(m,α, $\alpha_1,\ldots,\delta_1,\delta_2$) we use the probability P_α for that, the extreme algorithm gives absolutely correct classification, i.e. $P_\alpha = P\{Q(A^*) = 1\}$.

3. CLASS OF NORMAL TABLES GENERATED BY A(M,α, $\alpha_1,\ldots,\delta_1,\delta_2$), WITH CORRECT RECOGNITION.

Let us consider a subclass of normal tables set by additional conditions that satisfy the mathematical expectations of the components system of random variables η.

Let (T_1^a, T_2^b) be a pair of tables for training and control corresponding to the mathematical expectations $M\eta$. Let us consider the set of random variables Δ^* from $\eta \in \Delta$, such that, there exists an algorithm of ***A*** $\in$ ***A****(m,α, $\alpha_1,\ldots,\delta_1,\delta_2$)* which recognizes correctly the corresponding pair of tables (T_1^a, T_2^b). In such a case one can consider that the set Δ^* generates the class $T(\Delta^*)$ of normal tables correctly recognized by ***A****(m,α, $\alpha_1,\ldots,\delta_1,\delta_2$)*.

It is proved that if $\eta \in \Delta$, then $\sigma_{l_i}, \sigma'_{l_i} \to 0,\ l = 1,2,\ i = 1,2,\ldots,n$, exists a probability $P_\alpha \to 1$. In such a case for enough small dispersions of components of random variables $\eta \in \Delta^*$ the extreme algorithm $A^* \in$ ***A***(m,α, $\alpha_1,\ldots,\delta_1,\delta_2$) works correctly and the probability is close to 1. The results obtained for P_α are effective in case for enough small dispersion.

4. A CRITERION FOR OPTIMAL RECOGNITION

Proposition: The system of random variables $\eta \in \Delta$ belongs to Δ^* if and only if when there are non-negative numbers $\gamma_1, \gamma_2, \ldots, \gamma_n$ such that

$$R(a_s, b_s) < R(a_l, b_s) \quad s, l = 1,2, \quad s \neq l \tag{3}$$

where

$$a_s = (a_{s1}, a_{s2}, \ldots, a_{sn})$$
$$b_l = (b_{l1}, b_{l2}, \ldots, b_{ln}) \quad s, l = 1,2$$

$R(a_s, b_s)$ is the number of inequalities, which do not satisfy $\left|a_{s_i} - b_{l_i}\right| \le \gamma_i, \quad i = 1,2,\ldots,n$.

Proof: Let us assume that there exist $\gamma_1, \gamma_2, \ldots, \gamma_n \ge 0$, such that they satisfy the inequity (3). Let us define the algorithm ***A*** $\in$ ***A***$(m, \alpha, \alpha_1, \ldots, \delta_1, \delta_2)$, in the following way:

$$m = 1, \ \alpha = 0, \ \alpha_i = \gamma_i, i = 1,2\ldots n, \ \delta_1 \le m, \ \delta_2 \le \frac{1}{2} \qquad (4).$$

Let (T_1^a, T_2^b) be a pair of tables for training and control that corresponds to $\mathrm{M}\eta$, and the rows of T_2^b be recognized by the algorithm ***A.*** According to the algorithm defined in [2] we obtain:

$$\Gamma_l^{S'_r} = \sum_{q=(l-1)m+1}^{lm} \sum_{\lambda=0}^{r} C_{n-p}^{m-\lambda}(s_q, s_{r'}) C_p^{\lambda}(s_q, s_{r'}), \quad r = 1,2,\ldots 2t, \quad l = 1,2 \qquad (5),$$

where $p(s_q, s_r)$ is the number of non-met inequities $\left|a_{q_1} - b_{r_1}\right| \le a_i, \ i = 1,2,\ldots,n$,

$$s_q = (a_{q_1}, a_{q_2}, \ldots, a_{qn}), \quad s_{r'} = (b_{r1}, b_{r2}, \ldots, b_{rt}).$$

Since $\alpha_i = \gamma_i, \quad i = 1,2,\ldots,n$, then according to (4), we obtain

$$p(s_q, s_{r'}) = R(a_{q'}, b_{r''}) \qquad (6)$$

where $q' = \left[\frac{q-1}{m}\right] + 1, \quad r'' = \left[\frac{r-1}{t}\right] + 1, \quad 1 \le q \le 2m, \quad 1 \le r \le 2t$.

([***x***] is the integer part of ***x***).

By substituting (6) in (5) and taking into account (4), we obtain:

$$\Gamma_l^{s'_r} = m(n - R(a_l, b_{r''})), \quad r = 1,2,\ldots,2t, \quad l = 1,2 \qquad (7).$$

If $r \le t$ then from (7) and (3) it follows that:

1) $\Gamma_1^{s'_r} - \Gamma_2^{s'_r} = m(R(a_2, b_1) - R(a_1, b_1)) \ge m$,

2) $\dfrac{\Gamma_1^{s'_r}}{\Gamma_1^{s'_r} + \Gamma_2^{s'_r}} > \dfrac{1}{2}$

If $r \ge t+1$ then from (7) and (3) follows

1) $\Gamma_1^{s'_r} - \Gamma_2^{s'_r} = m(R(a_1, b_2) - R(a_2, b_2)) \ge m$,

2) $\dfrac{\Gamma_2^{s'_r}}{\Gamma_1^{s'_r} + \Gamma_2^{s'_r}} > \dfrac{1}{2}$.

In this way the algorithm ***A*** optimally recognizes all rows from T_2^b, therefore $\eta \in \Delta^*$.

Let $\eta \in \Delta^*$ i.e. be that there exists an algorithm ***A*** $\in$ ***A***$(m, \alpha, \alpha_1, \ldots, \delta_1, \delta_2)$ that correctly classifies the pair of tables (T_1^a, T_2^b) corresponding to $\mathrm{M}\eta$. Let $m^*, \alpha^*, \alpha^*{}_1, \ldots, \alpha^*{}_n, \delta_1^*, \delta_2^*$ be the values of the algorithm's parameters. Then from the decision rule ***A***$(m, \alpha, \alpha_1, \ldots, \delta_1, \delta_2)$ it follows that

$$\sum_{\lambda=0}^{\alpha}\binom{m-\lambda}{n-m}(a_l,b_l)\binom{\lambda}{p}(a_l,b_l)-\sum_{\lambda=0}^{\alpha}\binom{m-\lambda}{n-p}(a_s,b_l)\binom{\lambda}{p}(a_s,b_l)\geq\delta_1^*,\ s\neq l, s,l=1,2 \quad (8)$$

where $p(a_s,b_s)$ is the number of valid inequalities

$$|a_{li}-b_{si}|\leq\alpha_i^*,\quad l,s=1,2,\quad i=1,2,...,n.$$

Let $\gamma_i=\alpha_i^*,\quad l,s=1,2;\quad i=1,2,...,n$ …Then

$$R(a_s,b_s)=p(a_s,b_s)\quad l,s=1,2 \qquad (9)$$

Then from (8) and (9) it follows that

$$\sum_{\lambda=0}^{\alpha}\binom{m-\lambda}{n-R}(a_s,b_s)\binom{\lambda}{R}(a_s,b_s)>\sum_{\lambda=0}^{\alpha}\binom{m-\alpha}{n-R}(a_l,b_s)\binom{\lambda}{p}(a_l b_s)\quad s\neq l,\quad s,l=1,2 \text{ and}$$

$$R(a_s,b_s)<R(a_l,b_s)\quad s,l=1,2,\quad s\neq l.$$

Theorem: For the class $\mathrm{T}(\Delta^*)$ of normal tables, optimally recognized from the class of algorithms ***A****(m,α, α*~1~*,…,δ*~1~*,δ*~2~*)*, the probability $P_\alpha\to 1$, provided $\sigma_{l_i},\sigma'_{l_i}\to 0,\quad l=1,2;\quad i=1,2,...,n.$

Proof: According to the criterion for optimal recognition there are $\gamma_i\geq 0, i=1,...,n$, such that

$$R(a_s,b_s)<R(a_l,b_s)\quad s,l=1,2\quad s\neq l.$$

Then $R(a_l,b_s)>0\ s,l=1,2\quad s\neq l$ and therefore that

$$N_{ls}=\left\{i/|a_{l_i}-b_{l_i}|>\gamma_i\right\}\neq 0,\quad l\neq s,\quad l,s=1,2.$$

From the above follows:

$$\boldsymbol{min}\left(|a_{m_i}-b_{l_i}|-\gamma_i\right)=\varepsilon_0>0\quad m\neq l,\quad m,l=1,2;\quad i\in N_{kl} \qquad (12)$$

and for the respective pair of tables $(\mathrm{T}_1,\mathrm{T}_2)$, is proved that there exists an algorithm ***A****(m,α, α*~1~*,…,δ*~1~*,δ*~2~*)*, which works correctly.

Let us denote by $E(\alpha)$ the event:

$$E(\alpha)=\left\{|\eta_{qi}-\tilde{a}_{qi}|<\alpha,\quad |\eta'_{ri}-\tilde{b}_{ri}|<\alpha\right\},\quad q=1,2,\mathrm{K}\,2m,\quad r=1,2,\mathrm{K}\,,2t,\quad i=1,2,\mathrm{K}\,,n, \text{ where}$$

$$0<\alpha<\frac{1}{4}\alpha_0.$$

Let $T_1=\{S_1,S_2,\mathrm{K}\,,S_{m_l}\}$ and $T_2=\{S'_1,S'_2,\mathrm{K}\,,S'_{t_l}\}$ be tables including $S_q=(x_{q_1},x_{q_2},\mathrm{K}\,,x_{q_n}),\quad S'_r=(y_{r_1},y_{r_2},\mathrm{K}\,,y_{r_n}),\quad q=1,2,\mathrm{K}\,,2m,\quad r=1,2,\mathrm{K}\,,2t,$ and let the algorithm ***A*** have the following parameters [1]:

$$k=1,\ \alpha=0,\ \alpha_i=\gamma_i+\frac{1}{2}\varepsilon_0,\quad i=1,2,\mathrm{K}\,,n\,,\ \delta_1\leq m,\ \delta_2\leq\frac{1}{2} \qquad (12)$$

$$\Gamma_l^{S'_r}=\sum_{q=(l-1)m+1}^{lm}\left[n-p(S_q,S'_r)\right], \qquad (13)$$

where $p(S_q,S'_r)$ is the number of non-satisfy inequalities $|x_{q_i}-y_{r_i}|\leq\alpha_i,\quad i=1,2,...,n.$

Let $r \le t, \quad q \le m, \quad i \in N_{11}$, where $N_{ll}\left\{i / \left|a_{l_i} - b_{l_i}\right| \le \gamma_i\right\}, \quad l = 1,2$.

Then from (2) and (12) it follows:

$$\left|x_{q_i} - y_{r_i}\right| \le \left|x_{q_i} - \tilde{a}_{q_i}\right| + \left|y_{r_i} - \tilde{b}_{r_i}\right| + \left|a_{1_i} - b_{1_i}\right| < 2\varepsilon + \gamma_i < \frac{1}{2}\varepsilon_0 + \gamma_i = \alpha_i$$

and therefore

$$p\left(S_q, S'_r\right) \le R\left(a_1, b_1\right), \quad q \le m, \quad r \le t. \qquad (14)$$

Now for $r \le t, \quad q \ge m+1, \quad i \in N_{21}$, we obtain that

$$\left|x_{qi} - y_{ri}\right| \ge \left|a_{2i} - b_{1i}\right| - \left|\left(y_{ri} - \tilde{b}_{ri}\right) - \left(x_{qi} - \tilde{a}_{qi}\right)\right|.$$

By definition for the set N_{21} the inequality $\left|a_{2_i} - b_{1_i}\right| > \alpha_0 + \theta_i$ is true, and respectively $\left|x_{qi} - y_{ri}\right| \ge \varepsilon_0 + \gamma_i - 2\varepsilon > \gamma_i + \frac{1}{2}\varepsilon_0 = \alpha_i$ and therefore

$$p\left(S_q, S'_r\right) \le R\left(a_2, b_1\right), \quad q \ge m+1, \quad r \le t \qquad (15).$$

By analogy for $r \ge t+1, \quad q \ge m+1, \quad i \in N_{22}$, we obtain:

$$\left|x_{qi} - y_{ri}\right| \le \left|x_{qi} - \tilde{a}_{qi}\right| + \left|\left(y_{ri} - \tilde{b}_{ri}\right) + \left(a_{ri} - b_{ri}\right)\right| < 2\varepsilon + \gamma_i < \alpha_i$$

and therefore

$$p\left(S_q, S'_r\right) \le R\left(a_2, b_2\right), \quad q \ge m+1, \quad r \ge t+1 \ (16).$$

For $r \ge t+1, \quad q \le m, \quad i \in N_{12}$, we obtain

$$\left|x_{qi} - y_{ri}\right| \ge \left|a_{1i} - b_{ri}\right| - \left|\left(y_{ri} - \tilde{b}_{ri}\right) - \left(x_{qi} - \tilde{a}_{qi}\right)\right| \ge \varepsilon_0 + \gamma_i - 2\varepsilon > \alpha_i.$$

Consequently we have

$$p\left(S_q, S'_r\right) \ge R\left(a_1, b_2\right), \quad q \le m, \quad r \ge t+1 \qquad (17).$$

From the inequalities (10) – (17) it follows:

$$\Gamma_1^{S'_r} - \Gamma_2^{S'_r} \ge m, \quad r = 1,2,\ldots,t$$

$$\Gamma_2^{S'_r} - \Gamma_1^{S'_r} \ge m, \quad r = t+1, t+2, \ldots, 2t,$$

i.e. the rows of table T_2 are correctly recognized by the algorithm ***A***. It follows that $P\{E(\varepsilon)\} \le P\left\{Q\left(\tilde{A}^*\right) = 1\right\} = \tilde{P}$, but since the components of the random variables η, are independent, it follows that:

$$P\{E(\varepsilon)\} = \prod_{i=1}^{n}\prod_{q=1}^{2m} P\left\{\left|\eta_{q_i} - \tilde{a}_{q_i}\right| < \varepsilon\right\} \prod_{r=1}^{2t} P\left\{\left|\eta'_{r_i} - \tilde{b}_{r_i}\right| < \varepsilon\right\} \qquad (18)$$

using Tchebyshev's inequity we obtain:

$$P\left\{\left|\eta_{q_i} - \tilde{a}_{q_i}\right| < \varepsilon\right\} \le 1 - \frac{\tilde{\sigma}_{q_i}^2}{\varepsilon^2}, \quad q = 1,2,\ldots,2m. \qquad (19)$$

$$P\left\{\left|\eta_{r_i} - \tilde{b}_{r_i}\right| < \varepsilon\right\} \le 1 - \frac{\tilde{\sigma}'^2_{r_i}}{\varepsilon^2}, \quad r = 1,2,\ldots,2t, \quad i = 1,2,\ldots,n. \qquad (20)$$

From (2), (18), (19), (20) it follows that the probability $\mathrm{P}_\alpha \to 1$, provided that

$$\sigma_{l_i}, \sigma'_{l_i} \to 0, \ l = 1,2; \quad i = 1,2,\ldots,n.$$

Thus the theorem is proved.

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ОБРАЗОВАНИЕТО ЗА УСТОЙЧИВО РАЗВИТИЕ – ПОЖЕЛАНИЕ ИЛИ НЕОБХОДИМОСТ

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(Plenary report)

ВЪВЕДЕНИЕ

Втората половина на XX век и началото на XXI век са белязани от остри икономически, социални и екологични проблеми, засягащи все по-голяма част от територията и населението на Земята. Тези проблеми произтичат от човешката дейност – неразумно използване на природните суровини, нарастване числеността на населението, замърсяване на почвите, въздуха и Световния океан, унищожаване на дъждовните гори и кораловите рифове, изтребването на много видове организми, разрушаването на озоновия екран. В резултат започна глобално изменение на климата, нараства делът на пустинните и полупустинните области, появиха се нови болести и смъртоносни инфекции. Всичко това не просто влошава качеството на живот на хората, а застрашава бъдещото съществуване на човечеството. Кои са пътищата за предотвратяване на заплашващата ни глобална катастрофа? Образованието има ли отношение към тези проблеми?

ОСНОВНИ ПРИЧИНИ ЗА ГЛОБАЛНИТЕ ПРОБЛЕМИ

На едно от първите места сред причинителите на екологични проблеми в световен мащаб са *технологиите*, които се използват за извличане на природни ресурси и преработването им до краен продукт. Самите природни ресурси са разпределени неравномерно – по-голямата част от тях са на територията на слабо развити или развиващи се държави, използващи стари технологии и изпитващи финансов недостиг. Често безконтролно се използват изчерпаеми природни ресурси. Някои от ресурсите, които могат да се възобновяват, са част от сложни екосистеми и неразумната им консумация води до нарушаване или дори разрушаване на тези екосистеми. Например отделянето в атмосферата на флуорсъдържащи газове (фреони) води до разрушаване на озоновия слой, който е естествен щит срещу късовълновите ултравиолетови лъчи на Слънцето. Изчезването или изтъняването му ще изложи живите организми на вредното въздействие на тези лъчи, следствие от което е увеличаване на честотата на мутациите.

Друга причина за възникване и задълбочаване на проблеми е нерегулираното *нарастване на населението*. Това се отнася преди всичко за развиващите се страни, където са и най-тежките социални условия, свързани с бедността, недохранването и глада, липсата на удобни жилища, водоснабдяване и канализация, силно ограничения достъп до здравеопазване и образование.

Локалните *военни конфликти*, освен непосредствените жертви и разрушения, пораждат остри хуманитарни, социални, икономически и екологични проблеми в засегнатите държави.

СОЦИАЛНО-ИКОНОМИЧЕСКИ И ЕКОЛОГИЧНИ ПРОБЛЕМИ В БЪЛГАРИЯ

Нашата страна също е изправена пред екологични и свързани с тях социално-икономически проблеми [5, с. 39 – 42]. След 1990 г. в обществото ни започна цялостно социално-икономическо преустройство. Преминаването от централно планиране към пазарна икономика наруши макроикономическата стабилност на държавата. Задълбочиха се неблагоприятните демографски тенденции за високи темпове на външната и вътрешната миграция на населението. Наруши се стабилността на финансовия сектор. Забави се развитието на научния потенциал и високите технологии. В страната ни и преди съществуваха различни проблеми, но в новите условия те силно се изостриха. Влошеното качество на компонентите на околната среда у нас може да се илюстрира със няколко основни примери.

Водните ресурси са в недостатъчно количество и недобро качество. В същото време те са и неравномерно разпределени. Използването им и в настоящия момент е неефективно. По данни към 2003 г. едва 81 % от селата са водоснабдени с питейна вода. Загубите на питейна вода средно за страната са около 53 %. Канализационните мрежи са слабо изградени – делът на канализираните градове е 70,2 %, а на селата едва 2,1 %. В експлоатация са 61 съоръжения за пречистване на отпадъчни води, обслужващи 35,7 % от населението. В останалите населени места се използват септични ями и попивни кладенци, което води до замърсяване на подпочвените води.

Автомобилният транспорт причинява значително замърсяване на *въздуха* в населените места и особено в големите градове. Това се засилва от транзитното преминаване на автомобилни потоци през тях. Регистрирани са места с трайно присъстващи наднормени концентрации на вредни вещества във въздуха.

Значителна част от *земите* в страната ни потенциално са застрашени от ерозионни процеси. Не малко терени са замърсени с тежки метали.

България е сред богатите на *биологични видове* европейски страни. Защитените територии съставляват 5 % от площта на страната, но управлението им е недостатъчно ефективно. Горите заемат около 35 % от територията на България, но състоянието на горските фондове се влошава. Например масово е заместването на разнообразните естествени местни гори с хомогенни гори от видове с висока икономическа изгода (например хибридни тополи в заливните гори по островите на Дунав и край други големи реки). Делът на старите гори е едва 10 %. Нараства броят на застрашените от изчезване растителни и животински видове. Намаляват популациите на ловните видове.

Застрашително се разрастват проблемите свързани с *управлението на отпадъците.* От съществуващите в България 663 депа за отпадъци само няколко са изградени съобразно изискванията на Европейския съюз.

Нерешени у нас са и проблемите с *трангранчните замърсявания и шумовото замърсяване в населените места*.

УСТОЙЧИВОТО РАЗВИТИЕ – СЪВРЕМЕНЕН МОДЕЛ ЗА ПРЕОДОЛЯВАНЕ НА ГЛОБАЛНИТЕ ПРОБЛЕМИ

Изследването на причините, пораждащи опасността от глобална екологична катастрофа насочиха световното внимание към търсене на нови икономически модели за развитие на човешкото общество. С това вече са ангажирани много световни организации и институции, като например Организацията на обединените нации с нейната Програма за околна среда и Комисия по устойчиво развитие, Световната комисия по околна среда и развитие и много други. България също не остава встрани. Страната ни активно участва в множеството международни форуми и изяви и изпълнява поетите ангажименти.

Моделът на световно общество, в което икономическият напредък е свързан с висок жизнен стандарт и чиста околна среда вече е факт. Този модел доби популярност под наименованието *„устойчиво развитие”*.

По *определение* на Световната комисия по околна среда устойчиво е това развитие, което задоволява нуждите на настоящето без да застрашава възможността бъдещите поколения да удовлетворяват собствените си нужди (1987 г., доклад „Нашето общо бъдеще”).

В българското законодателство понятието „устойчиво развитие” е дефинирано по следния начин: „Устойчиво развитие” е развитие, което отговаря на нуждите на настоящето, без да ограничава и нарушава способността и възможността на бъдещите поколения да посрещат своите собствени потребности. Устойчивото развитие обединява два основни стремежа на обществото: а) постигане на икономическо развитие, осигуряващо нарастващ жизнен стандарт; б) опазване и подобряване на околната среда сега и в бъдеще” [4]. Следователно устойчивото развитие включва в интегрирана форма три компонента – икономически ръст, социален прогрес и опазване на околната среда.

Главните цели и основните изисквания за постигане на устойчиво развитие са изкореняване на бедността, изменение на моделите на потребление и производство, опазване и рационално използване на околната среда в интерес на социално-икономическото развитие. Положителни предпоставки за изпълнението на тези цели в световен мащаб са бързата интеграция на пазарите, движението на капиталите, увеличаването на инвестициите. В България постигането на устойчиво развитие е въведено като основен мотив в Управленската програма „Хората са богатството на България”, приета от правителството през 2001 г. Изпълнението на целите на устойчивото развитие предполага обединяване на усилията на цялото общество в национален и в световен мащаб.

Всеизвестна е ключовата роля на образованието за развитието на всички сфери, свързани с човешката дейност. В този смисъл, *за утвърждаването на устойчиво развитие, е необходимо и адекватно образование*. Образованието за устойчиво развитие не е нова тема [1, с. 5 – 15; 2], но дали все още е пожелание или реално съществува?

КОНЦЕПЦИЯ ЗА ОБРАЗОВАНИЕТО ЗА УСТОЙЧИВО РАЗВИТИЕ

Образованието за устойчиво развитие е процес, който трябва да съпътства целия живот на човека, от ранното детство, през основното, средното и висшето образование, включително и образование на възрастни хора. Следователно образованието за *устойчиво развитие излиза извън границите на формалното образование* и в никакъв случай не обхваща единствено училищното обучение.

Основните цели на образованието за устойчиво развитие са:

- да разкрие взаимодействията между икономическите, социалните и екологичните процеси и явления;
- да осигури критично отношение и по-голяма информираност за икономическите, социалните и екологичните проблеми в тяхното единство;
- да поощрява уважение и разбиране на различните култури и да приема техните приноси;
- да стимулира хората от всички възрасти да поемат своята отговорност за създаването на устойчиво бъдеще;
- да съдейства за формиране на гражданско общество.

В образованието за устойчиво развитие *централно място заемат въпросите* за равенство, солидарност и взаимозависимост в рамките на сегашното поколение и между различните поколения, за взаимоотношенията между богати и бедни, за взаимовръзките на човека и природата и отговорностите му към себе си, към човешкото общество и към околната среда. *Основен набор понятия* са: мир, демокрация, сигурност, човешки права, гражданство, бедност, социално и икономическо развитие, здравеопазване, равенство на половете, различия в културно отношение, опазване на околната среда, управление на природните ресурси, производство и консумация и др.

Очевидно е, че възрастовият обхват, целите и основните въпроси и понятия предполагат задължително *интегриран подход*, а не обособяване на учебен предмет, дисциплина, курс. Образованието за устойчиво развитие трябва да прониква във всички програми за обучение, на всички нива, включително професионалното образование, обучението на учителите и преподавателите, продължаващото през целия живот обучение на професионалистите и на хората, взимащи решения. За да се осъществят целите на образованието за устойчиво развитие, в процеса на обучение трябва да се прилагат *интерактивни технологии*, обучаваните да работят в екип помежду си и с обучителите. Важно място следва да бъде отредено на *неформалното самостоятелно обучение*, тъй като то е най-пряко ориентирано към обучаващите се и изисква активното им участие. Широки възможности предоставят и съвременните информационни и комуникационни технологии.

В самата същност на устойчивото развитие *като основа е заложена демократичността*. От тук следва, че образованието за устойчиво развитие може да е успешно, единствено ако е на същата основа. Процесът не трябва да се свежда до излагане на данни, нито само до търсене на решения, а до демократичен и перманентен дебат за състоянието, анализа на данните и възможните решения, съобразно конкретните условия.

ПРИРОДНИТЕ НАУКИ И ОБРАЗОВАНИЕТО ЗА УСТОЙЧИВО РАЗВИТИЕ

Природните науки са изключително благодатна почва за усвояване на знания, умения и отношения, свързани с устойчивото развитие. Това произтича от факта, че екологичният компонент е базисен, а екологията по своя генезис и обект на изследване е природна наука.

В настоящия момент в средното училищно образование в България природните науки са представени с предметите от културно-образователната област „Природни науки и екология”. Тези предмети са „Човекът и природата”, „Физика и астрономия”, „Химия и опазване на околната среда” и „Биология и здравно образование”. Изучават се в рамките на задължителната подготовка от III до X клас, а в рамките на задължителноизбираемата подготовка и в XI и XII клас. Учебното съдържание е избрано и структурирано въз основа на държавен стандарт [3]. Включени са основни ядра на учебно съдържание, обвързани със същността на устойчивото развитие. Такива са ядрата „Природни процеси и явления” и “Човекът и неговото здраве” в учебния предмет “Човекът и природата” и ядрата „Организъм - среда” и „Биосфера” в учебния предмет „Биология и здравно образование”. Формулирани са знания, умения и отношения, които трябва да бъдат усвоени от учениците към завършването на съответния етап или степен (приложение № 1). В учебните програми за различните учебни предмети и класове тези стандарти са декомпозирани в цели на обучение по съответните теми.

Явен акцент върху образованието за устойчиво развитие безспорно е поставен и в стандартите и учебните програми на културнообразователната област „Обществени науки и гражданско образование” (в учебните предмети “Човекът и обществото, “История и цивилизация”, “География и икономика”), както и в предметния цикъл „Философия”.

ЗАКЛЮЧЕНИЕ

Глобалните икономически, социални и екологични проблеми на настоящето са факт – и в световен, и в национален мащаб. Те влошават качеството на живот на хората и застрашават бъдещото съществуване на човечеството.

Икономическият напредък, свързан с висок жизнен стандарт и чиста околна среда може да се осигури чрез устойчиво развитие на обществото – развитие, благодарение на което и бъдещите поколения ще живеят поне толкова добре, колкото и ние.

Постигането на устойчиво развитие е едно от най-големите предизвикателства пред човечеството през XXI век. Основен принос за постигането му е отреден на образованието. Образованието за устойчиво развитие не е само пожелание – то вече е факт. До реалното му въздействие обаче има още дълъг път, който трябва да се измине и от природните науки.

Приложение № 1

Примери за знания, умения и отношения,
ориентирани към образование за устойчиво развитие
в Държавното образователно изискване за учебно съдържание

В резултат от обучението си ученикът:

- назовава дейности на човека, водещи до нарушаване на равновесието в природата (“Човекът и природата”);
- оценява влиянието на човека върху природата и причините за нарушаване на екологичното равновесие; предвижда резултати от промяна на екологичните фактори и въздействието на човека върху екологичното равновесие („Биология и здравно образование”);
- аргументира необходимостта от разумно използване на природните ресурси; доказва необходимостта от вторична употреба на материалите, безотпадъчни и безвредни производства („Химия и опазване на околната среда”);

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ТЕОРЕМИТЕ И РОЛЯТА ИМ ЗА РАЗВИТИЕТО ИНТЕЛЕКТА НА ЧОВЕКА

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(Plenary report)

Анотация

Като се използват от една страна факти от историята на математиката, а от друга – особености на теоремите и техните доказателства като специфични средства за фиксиране на математически знания, в доклада се аргументира твърдението, че "теоремите в математиката играят важна роля за развитието интелекта на човека, който ги изучава".

Ще си позволя да започна доклада си с една може би крайна констатация, която за някои колеги ще се стори и пресилена.

През последните 20 – 30 години постепенно се смениха два подхода към изучаването на теоремите по математика в българското училище. Преди в обучението по математика и особено в геометрията, централно място заемаше доказването на теоремите, разглеждани като основно средство за фиксиране на математически знания. Затова в уроците за нови знания почти всички изучавани теореми се доказваха, след това се изискваше учениците да заучават в къщи доказателствата. В следващите часове те първо бяха изпитвани да възпроизвеждат доказателствата, а след това се провеждаха упражнения за решаване на задачи с приложение предимно на новите теореми. Болшинството учители обикновено изпитваха освен за доказателства на новите теореми и за доказателствата на някои преди това изучавани теореми. Беше се наложила презумпцията, че всеки ученик трябва да може да възпроизвежда доказателствата на всички теореми, изучени от началото на учебния срок, до съответния момент на изпитването. По подобен начин и по другите учебни предмети, освен за новите знания, изпитваха и за всички "стари" знания от началото на учебния срок до съответния момент на изпитването. Положението по математика за учениците с по-малки възможности за усвояване на математически знания се усложняваше от още едно обстоятелство. От тях се искаше също да заучават и възпроизвеждат и най-сложните доказателства, каквито бяха например изводите на формулите за обем на пирамида, за обем на сфера, доказателството на теоремата за перпендикулярност на права и равнина и други. Ако се случеше ученици с малки възможности да бъдат изпитвани за тези доказателства, които естествено не бяха по силите им, те получаваха слаби оценки. Всички тези доказателства се изискваше да могат да бъдат възпроизвеждани и на зрелостния изпит. Не се допускаше официално никаква диференциация и масата ученици се надяваха на случайността "да имат късмет да им се паднат съответните въпроси". Тази съвсем необмислена крайност естествено настройваше учениците и родителите им против математиката в училище. Това доведе до другата неразумна крайност "да се отмени зрелостният изпит по математика".

Този факт, а също някои други фактори, на които нямаме възможност да се спираме, доведоха до нови, според мен, неразумни действия вече от страна на учителите, а именно: не само да не се изисква да се заучават в къщи доказателствата на изучаваните теореми, но и да не се дават такива от тях в учебните часове. Сега обикновено учителите съобщават догматично новите теореми, включително и формули и след това започват да решават задачи чрез тяхното използване. Ще си позволя да съобщя накратко за три такива случая и за реакциите на учениците, от които ги научих. Това, което ще съобщя, ми беше разказано и показано в тетрадки за работа в клас.

1) Учителка съобщава "За произволен триъгълник важи теорема, наречена "косинусова", която за всяка от страните му се записва така:

$$a^2 = b^2 + c^2 - 2bc\cos\alpha$$

$$b^2 = a^2 + c^2 - 2ac\cos\beta$$

$$c^2 = a^2 + b^2 - 2ab\cos\gamma$$

Теоремата се нарича косинусова, защото в нея се съдържа функцията косинус."

2) Учителка в 9. клас във връзка с квадратната функция $y = ax^2 + bx + c$ дава наготово "ан блок" формулата за координатите на върха p на графиката ѝ. $p\left(-\frac{b}{2a}, -\frac{b^2 - 4ac}{4a^2}\right)$, за абсцисите на пресечните точки, ако такива има, на графиката на $y = ax^2 + bx + c$ с абсцисната ос и на ординатата на пресечната точка на графиката с ординатната ос, а след това започва да упражнява учениците да ги използват.

3) Учителка в 12. клас съобщава наготово на учениците, че:

"Ако $f'(x) < 0$ в интервала (a, b), то функцията $f(x)$ е намаляваща в този интервал.

Ако $f'(x_0) < 0$ и преди x_0 $f(x)$ е намаляваща, а след x_0 – растяща, то в x_0 $f(x)$ има локален минимум."

Ако $f'(x_0) = 0$ и преди x_0 $f(x)$ е растяща, а след x_0 е намаляваща, то в x_0 $f(x)$ има локален максимум.

След това започва да упражнява учениците да използват тези твърдения за да решават задачи.

И в трите случая учениците се оплакаха, че нищо не са разбрали в училище. Тогава аз постъпих по следния начин:

В първия случай, чрез решаване на задачи за трите частни случаи $\alpha = 30°$, $\alpha = 60°$ и $\alpha = 45°$, $b = 8$ см, $c = 12$ см и сравняване на получените решения, доведох ученика до положение сам да изведе формулата $a^2 = b^2 + c^2 - 2bc\cos\alpha$ и да възкликне: "Ах, колко просто и лесно било всичко, а нашата учителка постъпи така гадно, че нищо не разбрах."

Във втория и третия случай аргументирах и илюстрирах графично съответните твърдения, след което учениците реагираха по подобен начин, само не употребиха думата "гадно". С подобни ситуации съм се сблъсквал и друг път. Ясно е, че при пълното отказване на аргументирането на твърденията, за учениците, израснали в съвременни условия, има нещо, което се губи. Опитах се да потърся отговор на въпроса "Какво е то?" Естествено първо се насочих към психологията.

Там обаче не намерих задоволително обяснение на проблема за убеждаването на хората във верността на твърдения, за които непосредствените възприятия (опитът, експериментът) не дават отговор. Затова се насочих към изясняване функциите на теоремите в математиката.

През последните години все по-трайно се осъзнава убеждението, че чрез тях се представят свойства или признаци на обекти от обемите на понятията, които не се посочват в техните определения. При това изисква се тези свойства или признаци да се доказват, само като се използват закони от логиката и правила за извод, а също и определения, аксиоми или преди това доказани твърдения. Приемането и превръщането на това изискване във вътрешна потребност и убеждение, а от там и в поведение, поне в част от хората прави от теоремите и техните доказателства специфично познавателно средство, а в определен смисъл и възпитателно средство. Човек започва да вярва само на информация, която получава от наблюдения и експеримент, т.е. от непосредствени възприятия или разсъждения, извършени на базата на правила, доказани теореми или закони.

Във втория случай се казва, че той отива зад границите на непосредствените възприятия. С развитието на цивилизацията и разширяването на училищната система вярата в информацията на базата на разсъжденията, основани на научни знания, в това число и на теореми, се разширява. Конкретно, асоциациите, които се създават в мозъка на човека като отражение на връзката между условието на всяка теорема и заключението ѝ, стават надеждно средство за досещане, т.е. повишават евристичните му способности. Нека изрично подчертаем, че именно доказателството от своя страна е основното средство за показване връзката между верността на условието на всяка теорема и нейното заключение. Когато последното по сложност е над възможностите на учениците, съответната връзка може да се илюстрира експериментално или чрез сравнение и обобщение на частни случаи.

Казаното означава, че теоремите и техните доказателства разширяват и задълбочават интелектуалните способности на учащите се.

Друга особеност на теоремите е, че след като бъдат доказани, те се използват наготово, т.е. чрез тях се икономисват труд, време и сили на хората, пред които е извършено доказателството. В случая се постъпва както при всички обикновени практически дейности. Известно е например, че след като дърводелец си направи или купи тесла, той вече я използва наготово, а не си прави всеки път нова тесла, когато му потрябва такава.

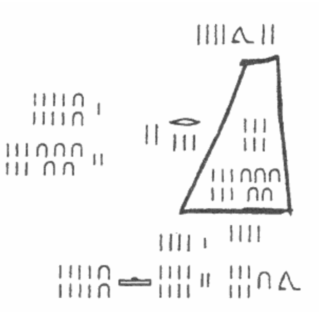
Посочените констатации, макар и верни, не ме задоволиха. В тях няма отговор на кардиналния въпрос: “Как хората се убеждават във верността на твърдения, за които непосредствените възприятия (опитът, експериментът) не дават някакъв задоволителен отговор?” Затова се обърнах и към историята. Тя, в резюмиран вид, ни дава информация:

В предгръцкия период на математиката се отговаря само на въпроса “Как?” При това почти навсякъде отговор на този въпрос се дава само за конкретни случай и то в алгоритмичен догматичен стил. Формулира се задача и веднага се излага решението ѝ. Ето два примера:

Задача “Намери обема на пирамида със страни на основата 2 и 4 лактя и височина 6 лактя”. (Има се предвид правилна пресечена четириъгълна пирамида).

Решение. “Събери заедно тези 16 с тези 8 и 4 и така ще получиш 28. изчисли $\frac{1}{3}$ от 6. така ще получиш 2. Умножи 28 по 2. ти ще получиш 56. виж той е равен на 56. ти го намери правилно.”

Този текст е написан с египетски йероглифи вдясно от чертежа (черт. 1а) с числа около него. На черт. 1б чертежът е представен със съвременни цифри, а текста сега записваме така:



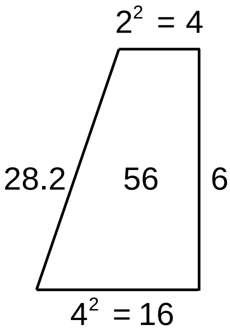
Черт. 1а

$$U = \frac{6}{3}\left(4^2 + 2.4 + 2^2\right) = 2\left(16 + 8 + 4\right) = 2.28 = 56$$

, т.е. като използваме формулата $U = \frac{h}{3}\left(a^2 + ab + b^2\right)$.

Задача 1. “Дължина и широчина на право поле. Дължината превишава широчината с 10. площта на полето е 11. Дължината и широчината колко са?”

“Решение. Раздели това, с което дължината превишава широчината на



Черт. 1б

половина, и ще получиш
повдигни го на квадрат)
големината на площта
това квадратен корен –
ще получиш 1
5 – ще получиш 11

Както се вижда тук
разликата на дължината
полученият резултат се
този квадрат се събира с
получения сбор се

Във връзка с
отбележа още два

5. вземи резултата 5 пъти (т.е.
– ще получиш 25. събери 25 с
11 – ще получиш 36. извлечи след
ще получиш 6. извади от шест 5 –
(широчината на полето). Събери 6 и
(дължината на полето).”

съвсем не се обосновава защо
и широчината се дели на две и
повдига на квадрат. Защо след това
“големината на полето” и от
намира квадратен корен и т.н.

решаването на задача 1 ще
факта:

1. Ако означим с p разликата на дължината и широчината на правоъгълника, а с q – неговото лице и следваме предписанията в решението, ще получим последователно:

$$\frac{p}{2};\ \left(\frac{p}{2}\right)^2;\ \left(\frac{p}{2}\right)^2 + q;\ \sqrt{\left(\frac{p}{2}\right)^2 + q};\ -\frac{p}{2} + \sqrt{\left(\frac{p}{2}\right)^2 + q};\ \frac{p}{2} + \sqrt{\left(\frac{p}{2}\right)^2 + q}.$$

Лесно се вижда, че последните два сбора са корени съответно на уравненията:

$y^2 + py - q = 0$ и $x^2 - px - q = 0$

и за нас сега това е съвсем естествено. Наистина, ако означим с x и y съответно дължината и широчината на правоъгълника, то според даденото в задачата:

$x - y = p$, $x.y = q$

от където получаваме уравненията

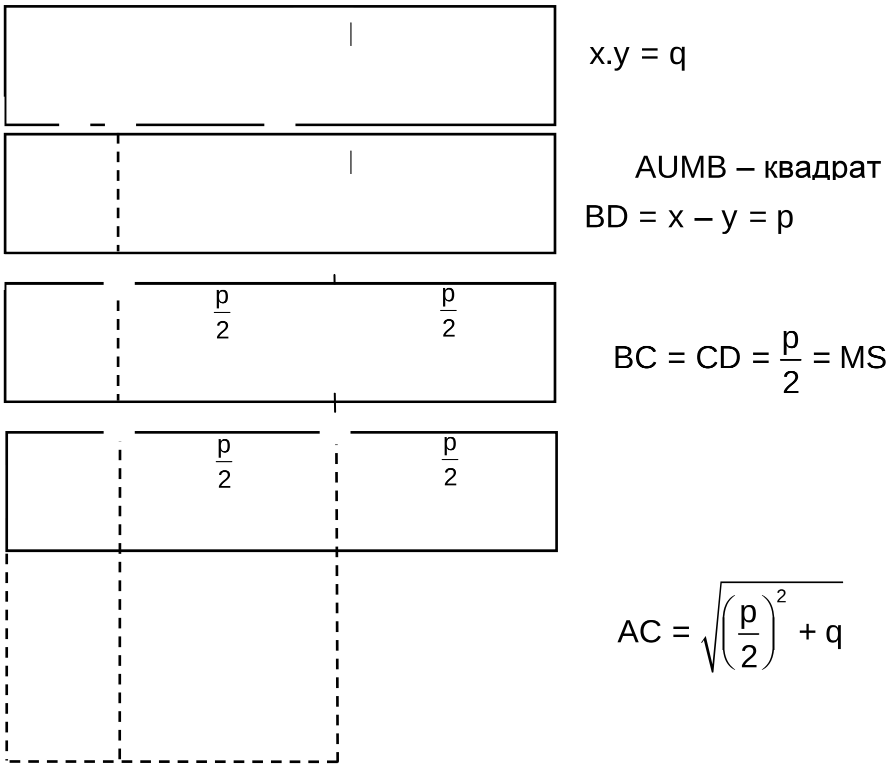
$y^2 + py - q = 0$ и $x^2 - px - q = 0$

Изложеното показва, че това, което вавилонците изчисляват стъпка по стъпка с конкретни числа, е равносилно на прилагането на съкратена формула за решаване на квадратни уравнения:

$y^2 + py - q = 0$ и $x^2 - px - q = 0$

Древните вавилонци обаче не са разполагали с буквена символика и не са могли по този начин да достигнат до предложеното от тях решение.

2. Като имаме предвид начина, по който са тълкували произведенията и втората степен чрез лица, при същите означения може да предполагаме, че те са извършили последователно следните построения (фиг. 2):



Фиг. 2

$\left(\frac{p}{2}\right)^2$ – лице на квадрата MNRS

$\left(\frac{p}{2}\right)^2 + q$ – лице на квадрата AKRC

$$y = AB = AC - BC = \sqrt{\left(\frac{p}{2}\right)^2 + q} - \frac{p}{2} = -\frac{p}{2} + \sqrt{\left(\frac{p}{2}\right)^2 + q},$$

$$x = AD = AC + CD = \sqrt{\left(\frac{p}{2}\right)^2 + q} + \frac{p}{2} = \frac{p}{2} + \sqrt{\left(\frac{p}{2}\right)^2 + q}.$$

Това е само една хипотеза, към която ни насочват, освен геометричните термини в задачата, още и геометричните начини за решаване на някои квадратни уравнения, използвани по-късно, както ще видим, от древните гърци. Тя, заедно с факта, че различни еднотипни задачи се решават по аналогични по между си начини, дава основание да предположим, че вместо днешната формула и заместването в нея, древният математик отначало е използвал като средство за убеждаване и умствена умора геометричният модел и аналогията, а след това постепенно е запомнял, че например за съответния тип задачи $BC = \frac{p}{2}$ е полуразлика на "дължината"и "широчината", а $AC = \left(\frac{p}{2}\right)^2 + q$ е полусборът на същите; освен това, ако от полусбора AC извадим полуразликата BC, ще получи широчината, а ако към полусбора AC прибави полуразликата CD, ще получи дължината.

Тези две задачи и техните решения са пример за това с какви средства и как се фиксират и представят математическите знания в предгръцкия период на математиката.

Известно е също така от историята, че през VI, V, IV и III век пр.н.е. в Древна Гърция и в обществената практика, и в математиката постепенно се налага "принципът за разумната обоснованост". Според него вярно е не това, което казват върховният вожд или върховният жрец, а онова, което може да се обоснове. Като последствие от този принцип в обществения живот постепенно се налагат "изборите", агитацията по време на изборите – въобще древногръцката демокрация, а в науката и особено в математиката – изискването да се отговаря на въпроса "Защо?" и от там – дедуктивното структуриране на знанията. Както се вижда в случая разликата между двата етапа в развитието на математиката съответства на два различни етапа, най-общо казано, в интелектуалното развитие на човечеството, а именно:

– етапът на наивното, безкритично и догматично вярване в неаргументирани твърдения на авторитетите;

– етапът на критичното приемане и аргументирано приемане на твърденията.

Преминаването от първия етап към втория в обществените отношения е траяло няколко века, но то е изиграло съществена роля в аргументирането на твърденията в науката и въобще в налагането на дедуктивното структуриране на научните знания и на математическите – конкретно. Това структуриране обаче, веднъж появило се, обратно е оказвало влияние на налагането в обществените отношения на изискването да се аргументира и доказва. Ярък израз на това влияние намираме в появата и налагането на логиката на Аристотел, а също и в логиката на стоиците. От тогава и до днес, особено след масовизирането на училищното обучение, аргументирането при излагането на научни знания формира в една или друга степен в ученическата възраст поне на част от хората това, което наричаме доказателствен стил на мислене. Тези хора са критични и обикновено приемат за верни само твърдения, установени опитно, с непълна индукция или дедуктивно.

От изложеното до тук не е трудно да се направи следният извод: Налагащата се у нас през последните 2 десетилетия тенденция масово да се съобщават само наготово (догматично) теореми, без да бъдат доказвани или поне опитно или индуктивно проверявани, означава отказване от възможността обучението по математика да бъде използвано като ефективно средство за интелектуалното развитие на българската младеж. Това означава още поставяне на обществото у нас по отношение на едни от най-силните фактори за интелектуално развитие в ситуация, близка до тази, която е съществувала за обикновените хора в Древен Вавилон и Древен Египет. Ясно е, че в такава ситуация у нас масово ще се формират граждани с незадоволителен доказателствен стил на мислене, т.е. с нисък интелект. А да се връщаме толкова назад през 21 век, едва ли е с нещо оправдано.

Къде е изходът?

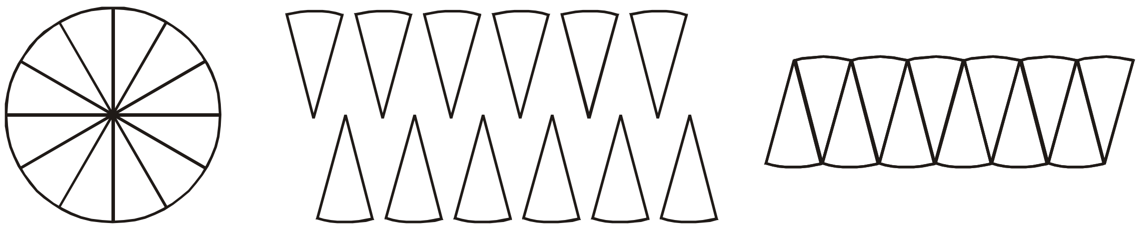
Изходът, както обикновено е при оформени два крайни и отричащи се подхода, и тук е някъде по средата. Донякъде той е подсказан от практиката на добрите учители и добрите преподаватели във В. У.

Например отдавна някои учители в пропедевтичния курс по геометрия убеждават учениците във верността на формулите чрез подходящи експерименти, като следните:

– за лице на кръг използват древноиндийския начин за нарязване на “два триона” на кръг от картон и получаване на фигура от вида на черт. 2;

Фиг. 2

– за обем на пирамида и обем на конус пълнене с пясък на подходящи съответни кухи призма и цилиндър с еднакви основи и височини;



– за лице на повърхнина на сфера, разрязване на същата на две полусфери и “покриване” на големия полукръг и самата сфера с навиване на подходяща връв и пр.

Други преподаватели вместо да изясняват някои твърдения в общия случай, което е сложно, разглеждат само частни случаи и доверявайки се на непълната индукция “достигат” до общия случай. Така се постъпва например при извода на формулата за общия член на аритметичната прогресия вместо да се използва методът на математичната индукция.

Акад. Н. Обрешков в лекцията си през 1954 год. по Висша алгебра за “Основната теорема на висшата алгебра” направо ни заяви, че доказателството е сложно и затова ще я използваме без да сме я доказали. Опитът да се оправдае сегашното пренебрегване на доказателствата на всички теореми, със случаи като този е неубедителен, най-малкото затова, че той е изключение. Освен това акад. Н. Обрешков явно и осъзнато посочваше защо постъпва така. По такъв начин фактически, макар и косвено, отново се внушава на учащите се, че нормалният подход е да се доказват теоремите.

Примери като посочените ни навеждат на идеята, че в обучението по математика е целесъобразно, когато математическите доказателства са сложни за съответната възраст на учащите се, те да не се дават, а да се заменят с експериментални дейности или с разсъждения на базата на непълни индукции, които все пак формират качеството критичност и усет за необходимост от аргументираност. По този начин се съдейства и за непринудено създаване асоциации между условията и заключенията на теоремите, на чиято основа се формира убеждението за тяхната вярност. Това означава, че освен възпитателни функции, те играят и важна евристична роля в обучението по математика.

Извод

Теоремите (включително и формулите) по математика и техните доказателства от древногръцкия период на математиката и до днес са важно средство за формиране интелекта на човека, защото съдействат за формирането на доказателствен стил на мислене. За убеждаване на учащите се в тяхната вярност е целесъобразно да се използват различни подходи, като експерименти, индуктивни разсъждения и дедуктивни разсъждения. Затова трябва да се преустанови вредната практика в училище само да се съобщават наготово почти всички изучавани теореми и след това те да се използват за решаване на задачи.

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ON THE NECESSITY OF LEARNING INFORMATICS BY PSYCHOLOGY STUDENTS

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Abstract. *This paper deals whit some aspects of the application of computer and information technology in the education of students in psychology. We presented the reasons, conditioning the necessity to study the discipline Informatics related to the contemporary requirements for the quality of students' education, as well as for their better realization as young specialists.*

The overwhelming development of computers and computer technologies during the last decade has resulted in their mass implementation in all spheres of life, including education. The knowledge of computer technologies has gradually becoming an indispensable part of the literacy of modern men, which is why it is natural that they are incorporated in the curricula of higher schools. Undoubtedly, any student should be familiar with the basic rules for operating a computer, its peripheral devices and the carriers of information. He or she should also be able to use programs for processing text, graphics, sound, tables, presentations as well as to be familiar with and know how to use the principle Internet services and applications.

This knowledge of computers and skills for their use are necessary not only for students in mathematical and technical majors, but also for those in all other fields. The aim of this article is to prove the necessity of the study of Informatics in the Psychology major in Higher University schools. To this purpose we will analyze the subjective attitude of students to the discipline on the one hand, and the objective necessity of its study on the other, which is imposed by the requirements of the educational degree, the major, the disciplines taught in it, and the instructors themselves.

According to the system of accrual and transfer of credits applied in universities [3] the acquisition of an educational degree requires a certain number of credits (240 for bachelor's and 300 for master's degree), which incorporate all the student's class and extracurricular activities, included in the curriculum (article 5). It is also required that the share of the extracurricular activities be not smaller than that of class activities (article 9).

The extracurricular activities of Psychology majors constitutes individual assignments like essays, mid-term and term papers, research projects, preparation and delivering of reports, etc. Part of these suggests the use of computers and computer technologies, which greatly relieves the work of students. With the preparation of others, computer technologies are indispensable and absolutely necessary.

Having in mind the nature of the Psychology major, we need to consider that in the course of their education the students in this major should prepare a significant number of individual projects. The major's curricular incorporates 32 disciplines (14 of which - optional) and in all these the student is obliged to hand in at least one written assignment (paper, research, report, term project). The use of computers for the finalizing of these assignments is stimulated by the instructors because they are not only aesthetically superior as compared to hand-written texts, but also facilitate the process of the correction of the papers, which are between 10 and 20 pages long depending on the requirements of the instructor.

For this reason, the work of students in psychology suggests ample use of text processing programs. The opportunities to typing, editing and processing a text with the help of a computer are employed with almost all individual assignments. The instructors themselves recommend that the papers, projects, and research be computer processed. Some instructors oblige their students to do so. This makes it difficult for a considerable part of them to hand in their assignments because they are not familiar with text processing programs. The anonymous poll of 36 third grade students in Psychology showed that two people cannot use a text processing program at all, while eleven others said that they found it difficult to do so.

These results were confirmed by a research into their written assignments (papers, term projects). 24 out of 70 papers were hand-written. Computer processed papers were often aesthetically flawed (40%) – incorrect alignment –23% out of the assignments, random tabulation –17%, non-standard margins – 7%, wrong numbering – 7%, etc. These flaws point to the insufficiency of students' knowledge of the ways of computer text processing and the inadequacy of their familiarization with the opportunities these give.

A number of research studies (tests, experiments, polls) should be made in the field of psychology in order to prove certain theories and hypotheses. Some specialized computer programs considerably facilitate the processing of the results. Unfortunately, only 18% of the informants can work with such products. This inadequate knowledge of the opportunities computer programs give makes students' education in certain disciplines very difficult. Such subjects are Experimental psychology, Psychological measurements, Psychology of Labor, Psychology of Deviant Behavior because psychological research is an indispensable part of these and the requirements of the instructors for the processing of the results suggest the use of computers. Some of them are even forced to spend some of the time in their classes teaching students how to use electronic tables or systems for data base management, which are exceptionally important in their work.

There are also specialized programs for statistical data processing. According to the curriculum, their study is optional. The increased efficiency of the mastering of the material in these disciplines stipulates that students have basic knowledge of computer science and computer technologies.

An important part of students' education is associated with their familiarization with programs for computer presentations. Computerized presentations give students an opportunity to be clearer, more persuasive and more attractive in their oral assignments. Unfortunately, students in psychology rarely use computers in their presentations because they are not familiar with the ways of preparing a computer presentation, neither do they have enough knowledge of the peripheral devices providing the presentation. Although instructors highly appreciate and give bonuses to such presentations, as well as students consider them more attractive and mentally stimulating, so far third grade students in Psychology have only delivered two presentations using this modern method.

The opportunity to use software for computer presentations, for computer text processing and computer processing of data facilitates and improves students' education. Undoubtedly however, Internet is among the guiding phenomena that make it easy for the students in psychology to complete their individual assignments. Nowadays, Internet is the most powerful source of information for humankind. Despite the variety of titles (1700 including 250 in electronic version) in the Psychology reading room at the library of South-West University "N. Rilski", it is quite recommendable to use Internet resources in search of information from all over the world. The options Internet gives also facilitate the work of international students studying in Bulgaria. Luckily, this information source is greatly used because of our students' sufficient knowledge of its potential and the ways to use it. According to a test for academic motivation, completed by 36 students in Psychology [1, 4] nearly 78% shared that often or very often they looked for information in Internet, while practically all of them were familiar with its informational capacity.

The aforementioned reasons determine the substantial role of computer technologies in the educational activities of students in Psychology. It comprises the opportunities for typing, processing and storing texts in a computer; for processing big bulks of data; for creating and working with computer presentations; for looking for information in Internet, which are necessary for the preparation and completion of individual assignments. Besides this direct application of computer technologies in the process of students' education, their knowledge in this field is closely associated with their future professional realization. A professionally significant quality of any psychologists, psychotherapists in particular, in connection with the nature of their job which suggests contacts with people of diverse social background, mindset, and education, is the variety of attitudes towards a large number of spheres of knowledge, including that of computer techniques and technologies.

Social and educational requirements nowadays impose a specialization on a larger scale for specialists in the field of Psychology. Their professional knowledge and skills are insufficient unless they are combined with knowledge and skills in other spheres of science and specifically in the use of modern technological means. The employer often demands that young specialists have knowledge and skill for operating a computer (computer literacy). The study of Informatics in higher schools which brings in a certain number of credits is noted down in the diploma and this can be considered a computer literacy document. What is more, a grade in a diploma from a prestigious university is much more reliable than a certificate from a course, sometimes of suspicious origin.

In spite of the substantial application of computers and computer technologies in the process of education of students of Psychology, this discipline is not part of the curriculum. Maybe they rely on the knowledge acquired in high schools. Indeed, over the past years the subject Information technologies has been averred as obligatory in high schools. In practice, however, not all schools have an adequate number of modern enough computers to carry out quality education and to familiarized students with the opportunities computer technologies give. Likewise, students at older ages (above 24) have not studied Informatics at high school, which makes it exceptionally difficult for them to continue their education in higher schools. This particularly refers to extramural students, who are by definition older.

These circumstances that refer to the level of students' familiarization with the opportunities computers and computer technologies give are confirmed by the poll carried out among 36 third grade regular students in Psychology. The results showed that over 1/3 of the informants have not studied Informatics at all in high school, which is why they have serious trouble using computers. It is indicative that those who have studied Informatics at school also testified to considerable gaps in their knowledge in the field, although they have mastered up certain theoretical facts and practical skills. Because of all, this part of the students rely on help from friends and acquaintances in completing their individual assignments, and some even shared that they have used the paid services of firms in order to type their papers and term projects.

Naturally, all this has a negative impact on students' satisfaction with the level and quality of their education. An anonymous survey with 109 university students [2] showed that the majority (73%) were pessimistic about the quality of their higher education. In terms of this, we should bear in mind the role of the computer in the enhancement of the quality of higher education. The tendency for computes to be used in the process of education has been quite dynamic recently. High schools and universities introduce computers and computer laboratories with state-of-the-are technology on a wide scale. The department of Psychology has recently acquired a modern psychological lab (2001) equipped with 8 computers. Access to the lab is free and students can use its modern equipment at any time. Unfortunately, the capacity of the lab is not used to its full degree (or is used to purposes different from those of education) because of the insufficient theoretical and practical preparation of students in the field of computer technologies.

The facts we have exposed and which are connected with the legal requirements of the educational degree and the major as well as those obtained through surveys are indicative of the significance of Informatics in the process of the education of students in Psychology. What is more, the implementation of the knowledge in the field of computer techniques and technologies surpasses the frames of the educational process and encroaches upon and influences positively the future professional performance of young psychologists. That's why the study of Informatics by students in Psychology is expedient and mandatory because it not only gives them an opportunity for a high-quality education, but also a perspective for a better professional realization.

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MAN AND NATURE, 5^{TH} FORM, CHEMISTRY MODULE TEACHER TRAINING COURSES

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Abstract. *Teacher chemistry training courses in Man and Nature curriculum subject for 5th form are discussed. The training modules are aimed at the easier adaptation of the available cadres and teachers' professional development. Authors share their observations and conclusions drawn on the work with about 150 teachers. That sort of short term course does not provide the necessary complete integral training but gives the possibility the three parts of the Man and Nature subject to be taught at primary school level. Special attention is paid to the visualization, demonstrations and experimental work. The integral approach to phenomena is shown by an example of hands-on activity related to chemical, biological, physical and ecological issues. Suggestions are proposed on the basis of the course participants' opinion and instructors' experience.*

Key words: *Science, Education, Chemistry, Teacher qualification, Experiments*

INTRODUCTION

The introduction of the new curriculum subject Man and Nature into fourth, fifth and sixth form curricula caused some expected difficulties for teachers, and most probably, for both students and parents. There are national shortages of teachers who are qualified in this integral subject, there are biology, physics and chemistry teachers who need further professional education and training as they work. Teacher training courses in separate modules are a good and useful initiative to prevent and overcome a part of the existing problems and to help teachers in their professional qualification to face the new challenges.

The authors of the present paper will share their observations about a number of chemistry training courses carried out in different Bulgarian regions. Some recommendations based on the analysis of the results and the conclusions drawn will be proposed in order to raise the efficiency of next training courses as well as for improvement and advancement as needed.

RESULTS AND DISCUSSION

The main conception of Man and the Nature school subject is to provide an introductory knowledge about nature, interrelating biological, physical and chemical fields of knowledge. This is a way to match our junior secondary education with that of the other European countries where traditionally the subject Science is taught.

The present Bulgarian Man and Nature programme of study [1] is divided into three modules, each focusing on a particular field of science: physics, chemistry and biology. For now, this three-part representation seems to be the most appropriate way to teach science at the junior secondary school level taking into account the easier adaptation of the available cadres ready to teach this triune subject. As a rule, Bulgarian science teachers are specialized in one or two subjects: physics, chemistry, biology only, or chemistry and physics, biology and chemistry etc. This situation requires training in an additional short-term module course targeted at the lacking qualification. That sort of course does not provide the necessary complete integral training but gives the possibility the three parts of the Man and Nature subject to be taught at junior secondary level. In future the professional training should be aimed at an integral rather than a differentiated in modules science content [2]. Students in this age group have to build on a general, integral prior knowledge about the natural and physical world instead of having separated ideas about biological, physical and chemical phenomena in nature. Consequently, the school programmes have to be redesigned and improved after that "intermediate stage". To make a summary, the chemistry module of the training course Man and Nature [3] addresses those teachers who have not been awarded the professional qualification "chemistry", "chemistry and physics", or "biology and chemistry". They are supposed to be not enough experienced and trained to teach chemical knowledge in fifth form. That is why the chemistry module programme concerns mainly specific chemical knowledge.

The content of this paper is based on observations on the teacher training courses carried out with about 150 teachers in the cities of Blagoevgrad, Pleven, Kranevo and Sofia. The participating teachers are graduated in biology, physics, physics and mathematics, biology and geography. Most of them teach physics, biology and natural history, some of them even chemistry, no matter they have not the corresponding qualification. About 10 instructors teach in mixed form classes. In all groups teachers took a positive and active position, and willingness to work. They showed ambitions to profit fully by lecturers' materials with a view to their future efficient use in classroom work.

During the training special attention is paid to visualization, demonstrations and experimental work. Most of the experiments can be carried out using available resources and on hand materials. In view of the lack of well-equipped chemical laboratories and facilities at the great number of schools it is important to be involved easy to perform experiments, which need simple chemical equipment. The experiments should be performed according to the safety rules in a way to avoid possible hazards and not to put in jeopardy students' health. A great number of experiments can be performed at home, for instance such as the study of the solubility, preparation and separation of mixtures, aqueous solution properties etc. Referring to both the theoretical and practical parts of the course, a special emphasis is placed onto students' own experience, to link it with the new topics of the science curriculum content, to apply a multidisciplinary approach in the context of ecological issues and preservation of the environment and health.

In the end of the Man and Nature chemistry module course the participants prepare and defend their own methodological projects. The latter are the best demonstration of the effect of the training. During the representation teachers shared a great deal of ideas involving the combination of different approaches, use of visualization tools, ICT and even out-of-class activities.

Along with the positive attainments, the design and the defending of the above-mentioned projects outlined some common for our education weaknesses. What particular inferences could be drawn from the projects presented?

Positive inferences:

- ✓ The teachers showed a good general pedagogical qualification.
- ✓ The teachers' learning of the necessary chemistry content and skills can be qualified as "well done".
- ✓ In most cases the projects have been implemented using modern methods of presentation.
- ✓ In a great deal of projects teachers proposed a variety of good practices, instructional strategies and approaches, some of them being hands-on and interactive.

Negative inferences:

- Most of the projects comprised too wide curriculum content.
- No reason or explanation was given why a particular instructional strategy had been chosen: a formal presentation of the goals, no detailed description of the expected outcomes. The link content ↔ instructional tools has not been clarified.
- Few problems aimed at the applying of new knowledge and skills by the students themselves during the current lesson or immediately after its completion.
- Summary lessons tend to repeat the concepts introduced before in a way which does not contribute to reveal the links between them and to develop them at higher cognitive levels.
- There is a trend to maximize the need of technical equipment without telling how this would contribute to more effective learning.
- Some presentations are designed using a great number of effects without any didactic ground.
- To be most effective ICT should be part of a planned experience and used by students collaboratively. When used less effectively, ICT is not interactive, providing a poor diet of PowerPoint presentations and virtual experiences which replace the first hand experiences needed by learners.
- The link: composition → structure → properties → application →

  environment preservation is underestimated**.**
- The possible interdisciplinary links, notably with physics and biology, are not implemented fully, so that a whole general idea for the phenomena in the world to be built up.

A general conclusion could be drawn on the basis of the observations and inferences above. Teachers need to engage students in laboratory investigations, so that to nurture curiosity about the natural world and include "hands-on, minds-on" inquiry-based science instruction. An example of such activity is presented by the American Chemical Society [4]. The experiment is easy to be done, pupils observe familiar phenomena from their every day lives. In this case, however, they perform a target observation aiming to make them to view a phenomenon from different standpoints: chemical, biological, and physical, and to relate it to ecological issues.

This activity is an example of the integral approach and gives children an opportunity to draw their own conclusions about the harmful effects on the environment and how they could be reduced and overcome.

EXPERIMENT: ***SIMULATION OF OIL SPILLS***

Key Concept: Oil spills or improper disposal of petroleum products may result in major environmental problems. There is also great potential for long-term environmental damage.

Skills: Observing, investigating, recording, hypothesizing

Objective: To simulate an oil spill and identify the characteristics of oil in water. To evaluate the effectiveness of various cleanup methods.

Content Focus: When we think of oil spills damaging the environment, we tend to think of largescale disasters. But non-point source pollution is probably more destructive. People dumping motor oil on the ground or in street sewers after changing the oil in their cars or ships dumping waste oil overboard occur much more frequently and are more difficult to clean up. When figuring out how to minimize damage to the environment as a result of an oil spill in the ocean, people must first consider the properties of oil and water. Some parts of the oil, fractions, are less dense than water and will float on the surface. Over time, the lighter fractions will evaporate and the heavier fractions will clump up and sink. Authorities also must consider whether the method of cleaning-up the oil spill will cause more environmental damage than another less effective method or even doing nothing. For example, burning an oil slick will remove it from the ocean water but it will pollute the air. There may be safer options. Some of these options include containment (keeping the oil together) with booms; adding dispersants (materials that break the oil slick up) such as surfactants; pick up with materials like straw or synthetic fibers; and treatment with oil consuming bacteria. An experimental method would require a special recovery boat to accompany every oil tanker. This boat would drive over the oil slick immediately following the disaster. The oil and some surrounding seawater would enter a large tank in the boat where the oil would float to the top and push the more dense seawater out of the bottom. If this method works on a large scale

— it has been tested with models — damage to the environment would be minimized and the recovered oil could be resold!

Advance Preparation:

• Divide students into groups.

• Prepare oil spills ahead of time by pouring water in a bowl for each group and adding 1 tablespoon of vegetable oil.

• Plan how you will cleanup from this activity. Each group will have a bowl of water with the oil spill and a bowl or other place to put the recovered oil, some water, and waste.

The oil spill activity could be discussed beforehand with the students so that each group can plan how they might try to clean up the oil spill. Either have them bring in materials to try or provide materials such as those included in the materials list. All groups do not need to try the same cleanup methods.

Materials (for each group): 2 bowls**,** 1 tablespoon vegetable oil**,** liquid detergent for dishwashers, paper towels**,** cotton balls**,** feathers**,** cotton string**,** nylon cloth - pieces about ½ cm wide by 5 cm long.

Procedure:

1. Using the various pickup materials your group selected, try to pick up the oil from the surface. Place the oil and your waste products in a separate bowl.

2. Compare your results with others who used different clean up tools.

Suggested Questions:

1. Why birds are affected by oil spills?

If you used feathers as one of your methods of cleanup, you may have an idea of one reason birds are terribly affected by oil spills. Why do you think getting oil on their feathers can be a life-threatening situation for seabirds? Birds spend much of their time taking oil from the preen gland near their tails and spreading it all over their feathers. This oil helps to keep the feathers flexible and waterproof and is part of how birds stay warm when it's cold outside. Since feathers absorb oil very well, a bird affected by an oil spill will absorb too much oil in its feathers, clogging the little hooks that keep feathers connected in an insulating coat around a bird's body. Cold water quickly rushes into the "coat" making the bird cold and heavy. When it tries to clean its feathers, it eats and smells oil, which is unhealthy. However, most affected birds die from hypothermia. People have tried washing birds in detergent to remove the oil and save them. NOTE: It is very important that birds be given a warm place to replace the natural protective oils on their feathers first. Without the oil, a bird's feathers do not form a waterproof coat around its body. So newly washed birds are at risk of hypothermia, too!

2. Describe what happens when oil is spilled on water.

The oil disperses over the entire surface of the water. Water is denser than oil. That is why the oil stays on the surface. Oil will not mix with water (isn't *miscible*). They are not miscible to any great extent.

3. What does the detergent do?

The detergent acts as a dispersant. It causes the oil and water to mix. The same effect is seen in washing greasy dishes.**CONCLUSION**

The results of training on Man and Nature, Chemistry module for 5^{th} form, show that for the present these courses seem to be an appropriate way to provide the needed specialists trained in the three basic subject fields of that curriculum subject. A summary of the opinion of the participants in chemistry module shows that this kind of training programmes is of benefit for teachers to extend their knowledge in prescriptive curriculum content and to update their pedagogical skills placing particular emphasis on modern teaching approaches and practices. The trainers should be specialists in methodology of science, who are to appraise both the learning of the subject science content and the adequate instruction, so as the State Core Curricula Requirements to be met. Furthermore, in the near future, students enrolled in university teacher education have to get integral professional training in Man and Nature in order to overcome the modular curricular design and resulting shortcomings in students' learning. Enquiry and hands-on activities are central to teaching and learning in science. Pupils should be taught to use exploration and investigation to acquire scientific knowledge and skills. They should be able to communicate and discuss what happened during their work, to make simple comparisons, and to try to explain what they found out relating their knowledge and understanding to domestic, personal health and environmental context. Through nurturing children' curiosity, developing their skills and increasing their understanding of the world around them, Science education in Man and Nature should engender an enthusiasm for learning that will stay with them throughout their lives.

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GRAPH THEORY AND DISCRETE OPTIMIZATION IN HIGH SCHOOL

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Abstract: *The authors offer a model of school curriculum and technology of education, which have taken into account the age peculiarities of school students. Pedagogical-psychological and methodological aspects and peculiarities of teaching mathematics have been considered in their relation to the offered model of education, which has been tested in several Bulgarian schools.*

The fulfilled experiment and the obtained results are extremely up-to-date in the context of the following idea:

- to humanize teaching mathematics through a rational rejection of the very theoretical and systematic approach, while emphasizing on the principals of clearness, accessibility, interdisciplinarity (integrativity), and applicability, which increase the interest and motivation of students.

- to assure a modern education in mathematics and realize a smoother transition between high school and higher education.

More than eight years ago together with our colleagues we started studying the possibility to use elements from the graph theory and mathematical programming in high school. Our purpose was: to offer the right model for a school curriculum; to adapt the respective topics for the students; to offer a technology of education, which could be tested; on the basis of the obtained results after the experiment to update the curriculum and the technology and to give an expert assessment of the possibility to use this model and technology. For this purpose we adapted the following topics from the graph theory and the discrete optimization for school students:

1. Introduction to graphs and networks. Some basic and essential notions, definitions and statements in graph theory – paths, cycles, types of graphs, matrix presentation of graphs, trees, Eulerian and Hamiltonian cycles, optimization problems in discrete structures.
2. Solving problems from the school course of mathematics with the help of graphs.
3. Optimal spanning trees in undirected graphs (maximum and minimum tree). Optimal tree structures in directed graphs.
4. Shortest paths algorithms. Dijkstra, Ford and Floyd algorithms.
5. Shortest paths algorithms applications. Maximum – capacity paths and the most reliable paths.
6. Maximum – flow problem. Minimum – cost flow. Dynamic flow.
7. Flow algorithms' applications – maximum cardinality, maximum - weight and perfect matching in bipartite graph.
8. Constructing project networks. Critical path method (CPM). Earliest and latest event time algorithms. Generalized project networks.
9. Eulerian graphs. Chinese postman problem. CPP algorithms.
10. Hamiltonian cycles. Traveling salesman problem. TSP algorithms.
11. Graph covering and colouring problems.

Some basic reasons for the inducement of this study are the following:

(1) "In the imagination of children throughout the world mathematicians are shaggy white hair nuisances with glasses and beards. This could be seen in a questionnaire, carried out by researchers from the University of Leicester in the United Kingdom and the University of Curtin, West Australia. The researchers asked children to draw a picture of a scientist. The more of the children made sketches of white hair teachers with eccentric hair-styles". Most children also stated that they would not like to become scientists because they thought that they would never enjoy. The survey, carried out among more than 4000 children, showed that children in other places throughout the world had similar notions.

(2) In a questionnaire carried out among school students in the 8-12 grades most of the respondents answered the question "What does mathematics basically deal with?" choosing one of the offered options "With proving theorems". When students answer the question "Is mathematics an applied science?", the most of them reply "Yes, mathematics is an applied science!". A very unpleasant conclusion automatically follows for those who teach mathematics: school children know that mathematics is an applied science but it is taught in a way that it very often looks boring, uninteresting, unexciting, and inapplicable science, which deals with proving theorems.

(3) Most of the people who deal with mathematics today have been taken by this science not through the classical textbook written for the school but through an interesting book in which great and deep mathematical problems have been presented in a human and understandable language with an emphasis on the ideas themselves, their development and applications. In Simon Singh's book „Fermat's last theorem" it is evident that Andrew Wiles, who proved it, has been won for the cause of mathematics when he was 11 and saw its simple formulation in an interesting book.

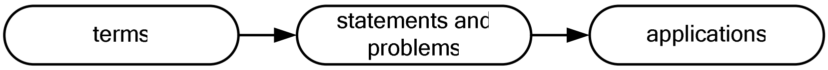
(4) In a number of questionnaires we found out that informatics is a favorite subject for the students, "not so much connected" to mathematics and unlike it much more interesting. This is an absolutely wrong idea, but probably it is grounded due to the way of teaching mathematics and the contents of the school curriculum.

(5) High school students showed a great interest when taught about the elements of graph theory and optimization. The fact that after several classes (for example Chinese postman problem and the traveling salesman problem) this interest remained and some of them carried on dealing with such problems themselves is very encouraging. And more, in the process of teaching we witnessed an increased interest by students who do not like mathematics.

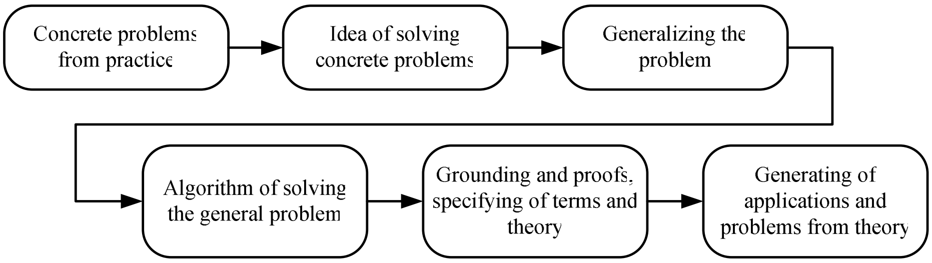
(6) The specifics of the particular high school are not taken into account. One and the same curriculum in mathematics is offered to all – there are no study groups, seminars, consultations.

The tendency in some countries is that the education in mathematics in non-specialized schools introduces mathematical ideas to students, not clumsy proofs. Mathematical formality and abstraction are really important and beautiful only to those who are more seriously interested in mathematics. In this type of schools mathematics cannot be only a purpose, but it needs to be a means, an instrumentarium for solving problems. It is difficult for an "ordinary" student to understand the beauty of the idea of continuity, when it is burdened with detailed proofs of many theorems.

(7) Psychologically it is clear that the receiving of real results in the process of one's own activity motivates the student to invest voluntary efforts to perform the activity. The classical model of teaching mathematics in our high schools is:



The basic idea of the study is to turn the classical model of teaching in the following way:



This approach is inductive in its essence. At the same time such an approach does have constructivist elements as well – knowledge is constructed by the people who learn when they are involved in active and effective learning, which assumes freedom of actions of the students and stimulating their critical reflection and ability to imagine verbally what they are doing. This approach is applicable to most of the topics of the curriculum.

(8) The massive introduction of computers and the very fast growth of the calculating possibilities of computers create natural conditions for solving optimization problems. The learning of problems of optimization in classes of mathematics and informatics gives extraordinary good possibilities for interdisciplinary connections between informatics, mathematics, physics, chemistry, biology, economics, etc.

The idea to teach students in the 8^{th} – 12^{th} grade topics traditionally aimed at universities looks exotic and very ambitious in the beginning. Later, when a good method and technology are found and they are mostly considering the age peculiarities of children, it may turn out that:

1. This is possible;
2. This is useful;
3. This is needed;
4. This is compulsory.

It is possible, because problems discussed in these spheres of mathematics and the ways of describing their decisions are natural and interesting. It is useful, because education in these spheres gives children a good instrumentarium for an easy solving of problems which look difficult and are a subject of study in a later course. It is needed, because they understand that mathematics is something interesting and something more general than calculation and solving equations and inequalities. It is compulsory, because it enriches the general knowledge of a person, his way of thinking, and makes the process of education continuous and interconnected from the elementary school till the graduation from university (and onwards).

We think that this suggestion has some advantages in an adequately projected educational environment.

There are good possibilities of differentiation of education in several directions:

a) profile differentiation – there is no division of curriculums and programs, which are different for the different profiles of high school education.
b) the model gives good possibilities for an internal (in a given profile) differentiation on the levels of abilities/inabilities. The assured possibility of every student to reach the goals of education in harmony with his abilities and the creation of suitable conditions to go to a higher level of knowledge and skills is a basic goal of each teacher.
c) The model offered allows making differentiations with regard to interests inside a given profile, e.g. the student chooses himself activities with regard to his interests, abilities, preferences, and possibilities for a future professional realization. Let us give a concrete example. While studying the topic of Flow algorithms, in the class we could identify three types of groups:

- students who use information technologies and program packages. These are people interested in solving concrete problems from real life with the use of a ready software and their basic activity would be modeling and getting familiar with the possibilities of the respective software;
- students with greater affinity to informatics and a desire to project and create software for solving a respective problem;
- students with affinity to mathematics and analysis, who are looking for the reasons of things, for mathematical validity of the given algorithms, looking for new and different ideas and algorithms for solving a problem.

d) Another type of differentiation is also possible – inside a given profile after the differentiation according to interests to make a differentiation according to possibilities.

There also exist very good possibilities to overcome the existing didactic contradictions inherent to the process of education in mathematics, which come out of its complexity – for example, one of the basic contradictions of teaching is connected to the process of outlining terms, e.g. the relation between abstract and concrete elements.

In the offered technology of education we start from L. S. Vigotsky's notion that there exist two levels of development of the individual – a level (zone) of the actual development and a level (zone) of the closest development. The selection of problems assumes both individual work of the student and collective work with the teacher. There are also problems whose complexity at the end becomes so overwhelming even when working with the help of the teacher. The emphasis in the concrete technology offered by us is connected to teaching in the zone of the closest development, which advances the maturing of the psychological functions in this zone and in a certain moment they gradually transfer to the zone of actual development. The sufficient number of problems for individual work is aimed at preventing, the disappearing, the loss of psychological functions from the zone of actual development. While giving new knowledge and skills or consolidating actual knowledge, the selection of problems assures enough individual work with knowledge on the demarcation line between the zone of actual development and of the closest development.

The grounds for the selection of this curriculum of study in summary they are mostly connected to:

- the possibility to make random selection of a part of the topics offered (for example topic 5 can be taught without knowing topics 2, 3, and 4, etc.);
- the possibility to teach without requiring additional knowledge from the students outside of the material provided in school;
- the possibility to establish a deeper connection and a smooth transition between high school and higher education;
- the possibility to establish intradisciplinary and interdisciplinary connections;

- the possibilities for a successful application of a differentiated approach (by profile, by interests, by abilities, etc.);
- the possibilities to overcome contradictions appearing in the process of teaching mathematics, which come out of its complex nature;
- the possibility to conform teaching with some classical results from genetic psychology, which are didactic rules;
- the possibility to use different methods, forms, and means of teaching and pedagogical influences;
- the possibility to consider the didactic principles of clearness, accessibility, individual approach in education, and as a result of this to assure stability of knowledge and an increase of the interest and motivation of students;
- the possibility to transform didactic principles into a means regulating and appointing the activities of teachers and students, e.g. similar to the possibility with the help of theorems to direct the solving of a given problem.

Each of the topics suggested has been developed in the following way.

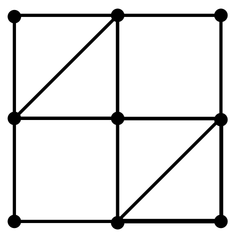
1. The goals and tasks have been formulated.
2. The basic knowledge needed for the topic and some of the key words have been given.
3. Knowledge and skills which students should acquire after introduction to the topic have been defined.
4. Some peculiarities of the curriculum, methodological instructions and notes have been given.
5. The recommended number of hours needed for the topic has been defined.
6. The necessary theoretical material and supplementary tasks have been given.
7. References suitable for the topic have been recommended.

At the moment the authors adapt some of the topics for students in the 5^{th} - 8^{th} grades. For example.

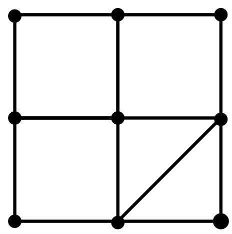
- Looking for Eulerian paths in undirected graphs.
- Looking for paths in labyrinths.
- Solving logical problems with the help of graphs.
- Solving problems in competitions and tournaments.
- Tasks to develop schedules.

For the purpose problems like the following have been used:

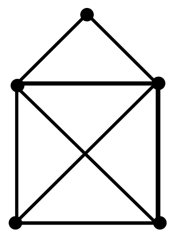
Problem 1. *Find an Eulerian paths in the following graphs (if it exists).*



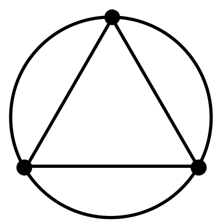
Plan 1.



Plan 2.



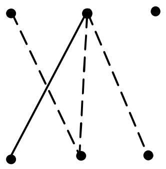
Plan 3.



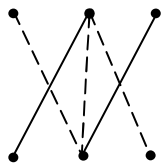
Plan 4.

Problem 2. *Three participants in an Olympiad in mathematics: Assia, Gallia, and Maya have solved a different number of problems – 3, 4, and 5. The problems solved by Assia are not 4. Gallia has solved fewer problems than Maya and they are an uneven number. How many problems has each girl solved? Which participant has scored best?*

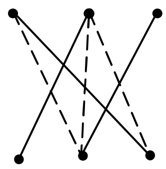
The graphic solution looks like this:



Plan 5.



Plan 6.

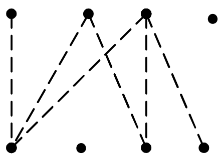


Plan 7.

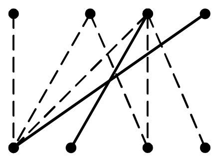
Problem 3. *Angel, Gosho, Mitko, and Tosho have the following family names: Vasilev, Ivanov, Petrov, Stoyanov. Find the family name of each of them, knowing that:*

a) Angel, Mitko and the student with the family name Vasilev study in different classes;
b) Mitko and the students with the family names Petrov and Stoyanov live in different streets;
c) Gosho and the students with the family names Vasilev and Petrov live in one and the same block of flats, but on different floors.

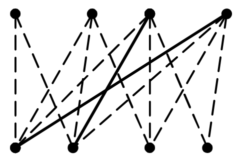
The graphic solution looks like this:



Plan 8.



Plan 9.



Plan 10.

Problem 4. *Sasho, Nasko, Velko, and Dimo are students in the third grade. Each of them plays on of the sports chess, gymnastics, soccer, and swimming. When the chess player participated in a tournament, Sasho and Nasko have been to the cinema. Velko and the swimmer go to a mathematics study group with the soccer player. The swimmer, Dimo, and Sasho go fishing together. Sasho and the soccer player sit at one and the same table in school. Guess who plays which sport?*

Response. Sasho – gymnastics, Velko – chess, Nasko – swimming, and Dimo – soccer.

Problem 5. *A basketball tournament was played by the teams of six schools, and each of them played against each other. How many games were played?*

Response. 15.

Problem 6. *A chess tournament was played by several girls, and each of them played against each other exactly once. How many are the girls, if the number of played chess games is 66?*

Problem 7. *In a kingdom there are 100 palaces and each of them is connected by road with 5 of the rest. Find the total number of roads connecting the palaces.*

Response. 250.

Problem 8. *There are 28 students in a class and they have exchanged pictures among themselves. Is it possible that 7 of them have exchanged pictures with exactly 5 of their fellow students, 13 – with exactly 6, and 8 – with exactly 3?*

Response. No.

Problem 9. (in a competition, 3rd grade). *Nedka, a third grade student, thought of a number. She added 2 to this number, and the obtained sum she divided to 3, then she multiplied by 4, subtracted 5, added 6, the result she divided to 7, added 8, and she obtained the smallest two-digit number with the same digits. What was the number, which Nedka thought of?*

Response. 13.

Problem 10. (2nd grade). *Several women and 23 men traveled in a bus. At the first stop 18 passengers got out of the bus, half of them men, and 12 persons boarded the bus, 5 of them women. Then the women became three times less than the men. How many women have traveled in the bus at the beginning?*

Response. 11.

Problem 11. *At the end of the first year a company tripled its capital, but it made expenses for 2 thousand leva. The next year the company doubled its available capital, but made expenses for 3 thousand leva. At the end of the second year it turned out that the capital of the company is 23 thousand leva. How much was the beginning capital of the company?*

Response. 5 thousand leva.

Problem 12. *A mother left peaches to her three sons and went to work. The oldest son woke up first, ate a peach, and divided the rest of the peaches in three equal shares, took his share and went out. Then the second son woke up, divided the peaches in three shares and took his share. The youngest son woke up last, divided the peaches in three shares, took his share, and 8 peaches left on the table. How many peaches has the mother left to her sons?*

Response. 28.

Problem 13. (about the devil). *A poor boatman was sitting on the shore of the river and he was thinking how to earn some money. The devil stood next to him and offered him: "I will double your money for every trip through the river but in exchange you will give me 24 stotinki each time." The boatman accepted the offer but for his surprise after the third trip he left without money. Define how much money did the boatman have in his pocket at the beginning?*

Response. 21 stotinki.

In the appendix are given several other problems like these.

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APPENDIX

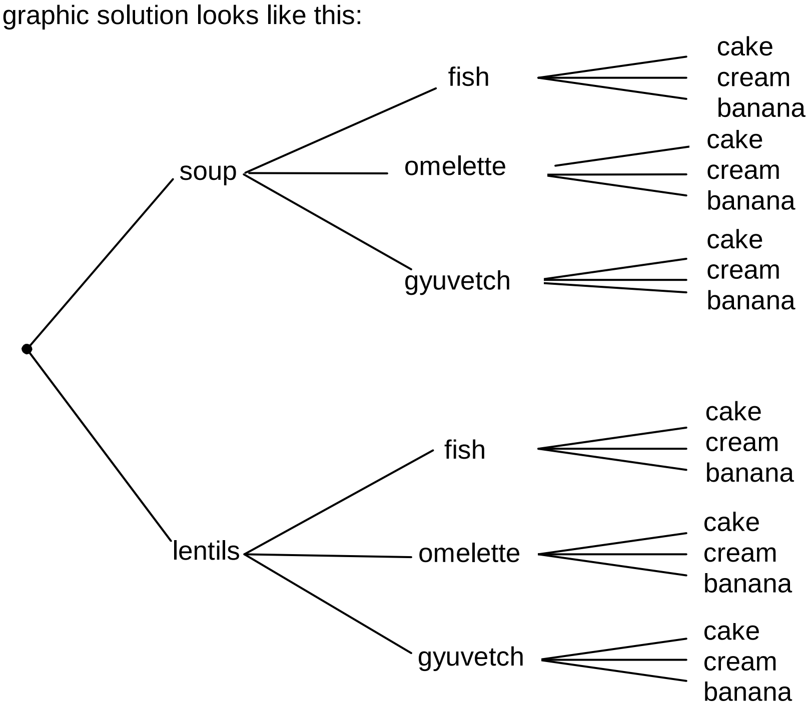
Problem 14. *In a student canteen one can choose exactly one from each of the three groups of meals in the offered list. Find the number of all possible combinations of meals?*

I
1. soup
2. lentils

II
1. fish
2. omelette
3. gyuvetch

III
1. cake
2. cream
3. banana

The graphic solution looks like this:



Plan 11.

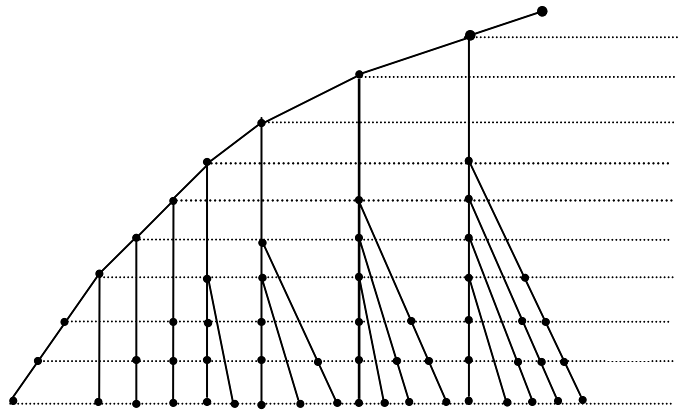
Response. 18

Problem 15. *How many four-digit numbers can be formed by the digits 0, 4, 6, and 7, so that in each given number each digit can be used exactly once?*

Response. 18

Problem 16. *A farmer had a young sheep, which gave birth for the first time and the lamb was female. Every following year the sheep gave birth to one female lamb. Each born lamb, after it became three years old, also gave birth to one female lamb, and after that it continued giving birth every year to one female lamb. How many lambs and sheep in total did the farmer have 9 years after the birth of the first lamb? (How the problem will change if every lamb gives birth for the first time when it is 2 years old?)*

The graphic solution looks like this:



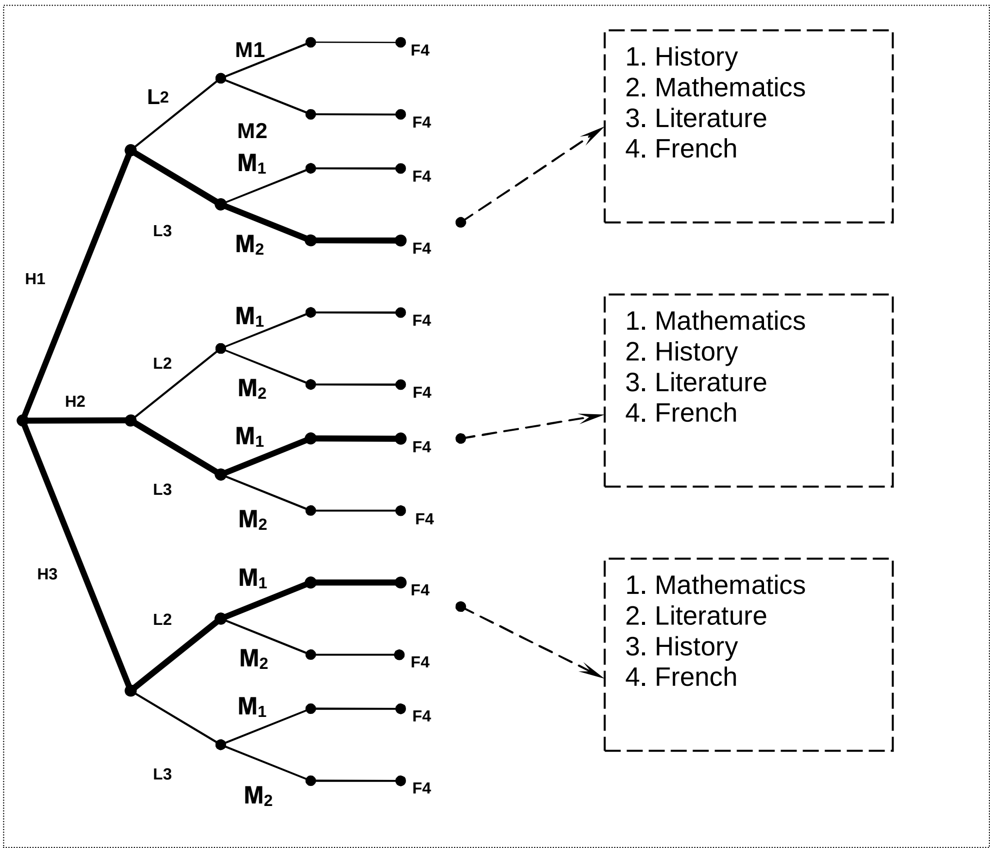
Plan 12.

Problem 17. *Write a syllabus of the classes on Monday, if the possibilities of the teachers in History, Literature, Mathematics, and French are known for this day:*

- The teacher in History can take the first, second, or third class.
- The teacher in Literature is available for the second and third class.
- The teacher in Mathematics can take the first or the second class.
- The teacher in French is available only for the fourth class.

How many versions of the syllabus can the school director write?

The graphic solution looks like this:



Plan 13.

MOTIVATION AS A PRINCIPLE IN TEACHING MATHEMATICS

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Abstract: *In this paper we proceed from a viewpoint that the most important goal in teaching mathematics is to develop students' skills for new knowledge based on creative and critical thinking, organization of the investigation work, exchanging opinions and discussions. For achieving this goal in teaching mathematics it is necessary to motivate students to learn mathematics, in other words to induce on the variety intellectual and other activities in the learning process.*

Keywords: *motivation, teaching mathematics, learning mathematics.*

1. INTRODUCTION

There are many factors that motivate students to learn mathematics, which can be integrated and synchronised together with different modern methodical approaches. Consequently, all this requires from the teacher: methodical variability and elasticity, respecting students' needs and capabilities for learning mathematics, as well as educating them in the spirit of ,,be own'', instead of ,,be like others''.

2. MOTIVATION AS A PRINCIPLE IN TEACHING MATHEMATICS

Motivating the student for learning, as one of the principles of modern maths teaching, is a system of factors that encourage the student for different intellectual and other activities in the process of learning. Besides emotional, motivation contains intellectual components as well. Thus, in the process of learning it is followed by emotions and a wish to overcome certain difficulties, that is, to solve the problem in the area of mathematics. Personal attitudes, interests, objective circumstances (hunger, exhaustion, material conditions for learning), the need for self-realization are all in close relation to process of motivating students learning.

The creative and inventive attitude is of the highest importance for learning mathematics. Such attitude towards learning develops intellectual capabilities to the maximum extend and helps the general positive development of the personality. It is developed by systematic practising of the habits for critical thinking.

Motivation has great importance for the success in learning. That means that, if the student is more involved in certain activity of learning, in that extend the achieved knowledge will be better in quantity and quality.

3. PLANNING MATHEMATICS CLASSES AND ITS INFLUENCE ON MOTIVATING STUDENTS

Teaching mathematics in class represents a series of directed and well-thought situations, and not spontaneous (usual) situations. That means that the motivation of students should be well prepared and previously thought.

Motivating students for learning mathematics represents influencing certain processes of establishing correlative relations between different students' needs and elements that should be taken into consideration by the teacher when planning the process of teaching mathematics.

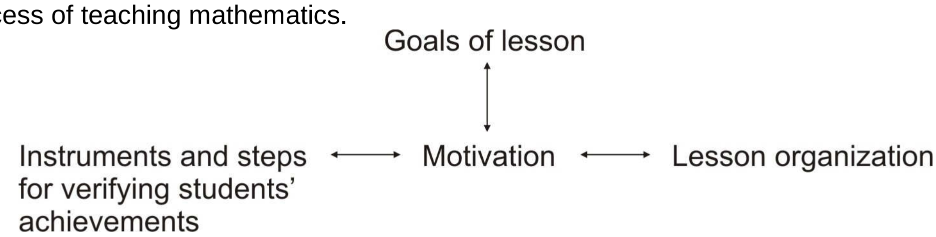

Fig 1: Elements of the lesson from the operational plan point of view

3.1. Announcing the goal

Creative teaching can be performed when creative activities are conducted, appropriate to the announced goals of the lesson. If the teacher tells the student how to perform each step and presents a copy of what the teacher is expecting from the students, than the students will not have a chance to solve the problems in creative way and independently. Every independent work allows every problem to be solved individually.

By using and developing cognitive motives, the teaching goals and exercises start to make sense for the student in a way that he/she adopts them, and activates his/her strength for their solving. Encouraged by the cognitive motives, the student is not satisfied with superficial, formal knowledge achievement. Motivated by the strive to know everything, he/she is trying to learn its essence, to discover internal relations, to set them in his/her system of knowledge, to verify them, to reactivate them on a continuous base, and similar. That means, that the existence and development of cognitive interests are important components in building conscious attitude towards learning, and the conscious and active solving of teaching topics and exercises.

Goal announcement on the class can be in the form of a sentence, a question or a problem. However, the best way for goal announcement on the class is when it is in a form of a problem, which can be reached by the students by developing the obtained theoretical knowledge through practical activity.

It should be taken into consideration that every problem and problem situation are not enough to motivate the students towards problem solving. It is necessary that the students accept the problem or the problem situation. With other words, the problem should be adopted as subjectively important goal, as external stimulus and condition that will influence the student's ratio. Only accepted and experienced problem situations are real problem situation.

To achieve that, the problem should meet the following criteria: to be real and convincing; not to overcome the level of the students' capabilities; to be clear and understandable in its formulation; to encourage certain ways of solving; to activate previous knowledge and experiences in relation to goals of the maths class; conclusions and findings of that particular problem to have implementation in new situations and similar.

The goal at the class can be announced differently, starting from the beginning of the class until the end. The most suitable moment is when the teacher feels that there is working atmosphere in the class, when students are concentrated and their psychological forces are activated towards future work. Unfortunately, we cannot determine which moment is the most favourable for goal announcement by using measure instruments. Thus, even if it is practically possible, then by measuring the working atmosphere, it can be easily destroyed. The teacher should announce the goal when he/she feels that there are favourable working conditions in the class.

Announcing the goal of the class gives clear perspective to students' work. Results coming from the goal announcement can be: students mobilize their own knowledge regarding the theme, interest for new learning is encouraged, and students are focused towards certain direction.

Such motivation aiming: creates and supports positive working conditions, produces and builds intellectual awareness for something new, unusual, special, provoke and develop cognitive interests, they put into force the senses and the intellectual capabilities, strengthen and refine the feelings of happiness, satisfaction that follow the acknowledgement, the finding and the discovery.

Example 1. Aim of the lesson: Noting a perimeter of a circle through its radius and vice versa.

In order to announce the aim of the lesson, the teacher poses a problem. Each group has three cans of a different size. They determine the perimeter of each of the cans by a non-strechable thread. They draw the bottom of each can on a piece of paper marked by millimetres, and next to it the length of the non-stretchable thread, the. the perimeter. As shown, in figure 1 and than they fill out the Tab. 1.

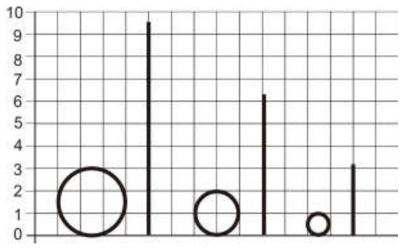

Fig. 2

	L	d	L/d
Can 1			
Can 2			
Can 3			

Tab.1

The teacher demands that the students make conclusions about what they see. If each group has performed well it determines that the requested amounts are approximately the same. That is the portion between the perimeter and the diameter of each circle is a constant. The teacher explains that this portion is usually marked with π. The number π is irrational number, but we are going to implement its approximately value of 3,14 in our calculation. This means that we can note the perimeter of a circle through its radius $L=2r\pi$.♦

3.2. Methods and instruments for following-up, checking and assessing students results

The system of follow-up, checking and assessing students results is the basis for objective presentation of changes and achievements not only in maths, but also in other subjects and also in general for the development of educational process. Follow-up, checking and evaluating achievements in maths comprise of many activities from the beginning of the student year until its end. The structure of this system can be seen through these questions: which activities should the teacher undertake at the class? By which methods and instruments? As well as from the answers of these questions. The teacher should be skilled and more engaged, should follow-up, record and evaluate students' results by using standardized tests, should explain students' results at each period of the student year, working results and each grade should have adequate objective background.

Grades as an indicator of the students' changes and results, no matter the form in which they are expressed, should be objective, the student should understand them, should be convinced that they are real, and they should indicate and motivate him towards further positive changes. Also, the negative grade, if it is objective and has background, and if the student understands and accepts the grade as such, can have a positive influence on motivation. If students have information for their achievements, they will know which actions should be undertaken in order to achieve better results.

In contemporary teaching, the work students perform is often based on the inner motivation. The work (learning) is pleasure. The students develop a sense that they can, that the results are a product of their effort, their self-confidence and self-assurance reinforces. The students cooperate and endure the work. They have no fear of being assessed, punished or being ridiculed for their failure.

Example 2. The teacher, in order to assess the knowledge for solving simple examples based on a numerical expression of the pupils in 5^{th} grade, on the beginning of the lesson chose the game I AM. Exercises are written on pieces of paper, for example: "I am 27, you are for 2 smaller than me", the pupil who has 25 answers, and the game continues until the pupil with the piece of paper on which END is written appears. During this game every pupil must be very careful and perform calculations all the time.♦

In traditional teaching, the outer motivation of the students prevails. Students work for a better mark or reward, or recognition for the learning and good behaviour, that is, not to be reprimanded or punished if they don't study. This is negative motivation whose main characteristic is fear of failure.

The teacher can compare the obtained results with the results from the beginning of the school year and by doing that the teacher will follow the improvement of the students. Besides the level of knowledge in mathematics, the teacher should also follow whether the student's behaviour improves; whether the student corrects his/her behaviour and avoids negative characteristics; whether the students respects the agreements and rules of conduct. During the lesson, the student's need for self-affirmation should be implemented. This motive is manifested, above all, through the individual's tendency to be recognised by the surroundings and have a certain position in the group. Therefore, during the lesson, praise for successfully completed work, approval, and consent for some idea should be implemented.

From all this we can conclude that during follow-up, checking and evaluating student's achievements, in function of the motivation of the students appear the following: praise, prize, acknowledgement, punishment, competition, cooperation, grade (mark), checking, sense for the experienced success, professional motives, etc.

3.3. Organization of the educational class

Teaching is a process in which teacher and students work together from one side and students work together from another. With a different choice and correct combination of teaching methods (further in the text learning methods), we can link both processes of teaching (teachers role) with the act of learning (students activity), as two links of the same process.

As long as the teacher knows many teaching techniques he/she can express creativity by correct choice and combination of techniques that will provide achieving the goals of each class, and at the same time, the curriculum will be more dynamic and more interesting for students.

Example 3. We consider the organization and activities for a part of a lesson. We can divide it into introductory part and main part.

In the introductory part teacher requires explanation for the procedure of graphic shift of a segment and constructing a segment equal to a pre-set segment. One of the students explains the procedure for graphic shift of line segments, and then every student individually constructs the assigned line segment. The teacher follows the work and corrects if necessary.

In the main part teacher gives instructions for reading the text for Graphical addition and subtraction of line segments. (Method INSERT or text method). Follows the work of the students while they are reading the text for Graphical addition and subtraction of line segments. With a coloured chalk, the teacher draws an INSERT table on the board. After the discussion, students fill out the table with what I know and what I don't know. The teacher explains the points that students don't know. The student read individually the text about Graphical addition and subtraction of line segments. Depending on their knowledge, they use the signs: √- what I know and ?- what I don't know which they write at the end of each paragraph. Then in pairs they clarify the parts they don't know. Every pair states what they know and don't know, having in mind not to have the same points with other pairs that have already said that.

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DOMINOES AND FRACTIONAL NUMBERS

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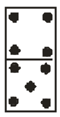
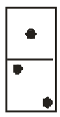
Abstract: *According to the Bulgarian Curriculum in Mathematics, fractions are studied in the 5th grade. Arithmetic operations with them are not attractive for pupils and lead to fatigue and repulsiveness. The paper proposes amusing elements to be involved in the process of teaching and learning. It turns out that the use of dominoes is suitable for the purpose. A didactical system of mathematical problems is developed applying dominoes as models of fractional numbers.*

To learn fractional numbers and arithmetic operations with them is an essential part of the aims in the Curriculum of Mathematics for the 5^{th} grade of the Bulgarian school. On the other hand, a long-lasting manipulation with fractional numbers is an obligatory activity in the process of forming of knowledge and skills but still leads to fatigue and to a decreased level of attention stability. This hinders fruitful learning itself. Not occasionally, in various textbooks in Psychology the activity under consideration is assumed as a suitable example of monotony, which hides a danger to cause ennui and discourage for those, who execute the activity. It seems to be possible to use dominoes as models of some fractions aiming at the elimination or at least the reduce of the negative effects of monotony and the resulting fatigue. The very fact, that the topic concerns a play, creates arguments for liberation from the eventually accumulated tension. It is well known that dominoes are suitable for demonstration of various mathematical relations, for the solutions of combinatorial and geometrical problems but in the present note they are applied to learn the topic "Fractional Numbers" namely. The main idea is that dominoes could be used in presenting common fractions and decimals. This could be done in the following way:

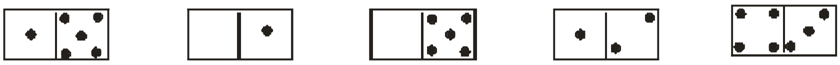
Common fractions

The separating line plays the role of a vinculum.

 $\frac{4}{5}$  $\frac{1}{2}$

Decimals

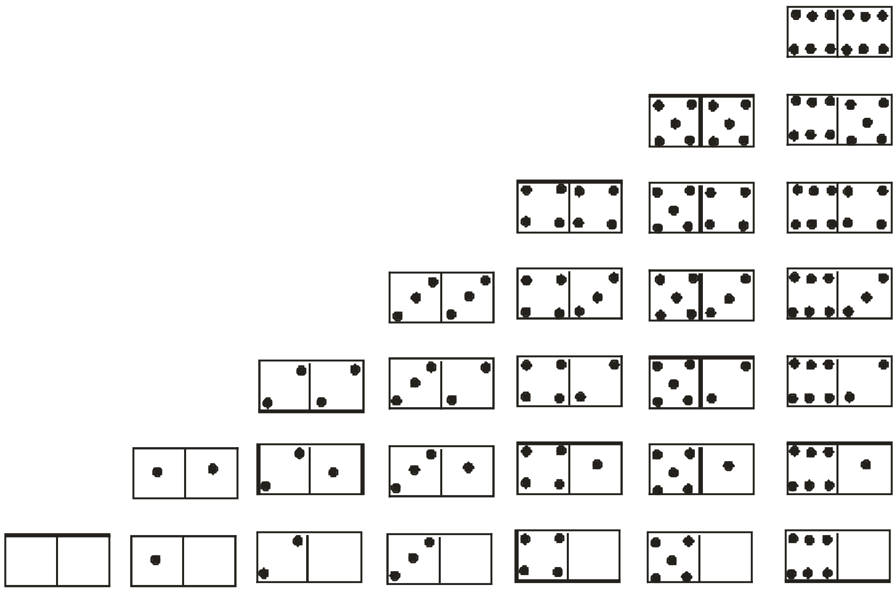
The separating line plays the role of a decimal comma.



1,5 0,1 0,5 1,2 4,3

The initial acquaintance with the suggested instruments includes an establishment of the number and the type of the dominoes from a full collection. This is a completely combinatorial problem, which could be solved by an exhaustion of variants respecting pupils' age. It is enough to start with the number 0 for example and to establish that this number could be combined with 0, 1, 2, 3, 4, 5 and 6 obtaining 7 dominoes in total. The number 1 comes next and it could be combined with 1, 2, 3, 4, 5 and 6 (6 dominoes in total). The turn is of the number 2, by means of which 5 new dominoes are obtained (with 2, 3, 4, 5 and 6), then the number 3 comes obtaining 4 new dominoes (with 3, 4, 5 and 6), by 4 – three dominoes (with 4, 5 and 6), by 5 – two dominoes (with 5 and 6) and finally by 6 – one domino (with 6). Thus, the total number of all dominoes in a full collection is equal to 7 + 6 + 5 + 4 + 3 + 2 + 1 = 28.

All the dominoes in a collection

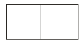


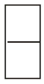
Some other combinatorial problems are proposed in the sequel (see problems 1, 2 and 3). Additional problems are possible too including ones which are connected with the rules of the play. We do not consider them. The basic aim of the paper is to present a didactical system of problems, which is directed to the improvement of the skills in manipulations with fractional numbers.

Problem 1. Are there dominoes by which it is not possible to obtain:

a) a common fraction; b) a decimal?

Solution: a) Taking in mind that a common fraction is well defined when its denominator is different from zero, we conclude that if a domino does not contain a zero then it represents two different common fractions. (Realizing that dominoes could be turned "up-down" is essential for the further problems too.) The domino with two zeroes is the only one by which no common fraction could be formed. Equally valuable common fractions could be obtained by all the others with a zero. In this case the corresponding value is zero.

b) Decimals could be formed by each domino. This time the domino  is not excluded and the decimal 0,0 corresponds to it (its value is equal to zero).

Answer: a) the domino  is the only one by which no common fraction could be formed;

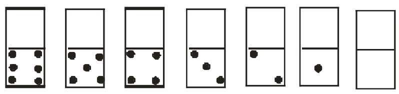
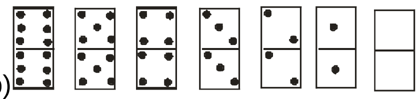
b) decimals could be formed by each domino.

Remark. The solution of the last problem is connected with the necessity of becoming conscious about the definitions of a common fraction and a decimal including the fact that the denominator of a common fraction is a number, which is different from zero.

Problem 2. By which dominoes is it not possible to obtain:

a) an improper common fraction; b) a proper common fraction?

Solution: The dominoes from a collection can be divided into three groups. They are in the first group those dominoes (in total 15), by which two different common fractions can be formed. For example, the domino with five and four points on it represents the common fractions $\frac{4}{5}$ and $\frac{5}{4}$. What is possible to do by each domino from the second group is to represent a unique common fraction. Such a domino contains only one zero and the number of the dominoes of this type is equal to 6. Also, they are in the same group the dominoes with one and the same number of points on their both sides. The number of such dominoes is equal to 6 too. It remains only one domino (third group) with no points on its both sides. As mentioned in the previous problem, this domino is special with respect to the formation of common fractions. Further, it is enough to remind the definitions of a proper common fraction and of an improper one: a common fraction is called to be proper if its nominator is less than its denominator; if the nominator is greater or equal to the denominator then the corresponding common fraction is called to be improper.

Answer: a)  ; b) .

Problem 3. Find the number of all different common fractions which could be obtained by a full collection of dominoes. How many of them are proper and now many improper?

Answer: 24 (the digits 0, 1, 2, 3, 4, 5 and 6; the common fractions $\frac{1}{2}, \frac{1}{3}, \frac{1}{4}, \frac{1}{5}, \frac{1}{6}, \frac{2}{3}, \frac{2}{5}, \frac{3}{2}, \frac{3}{4}, \frac{3}{5}, \frac{4}{3}, \frac{4}{5}, \frac{5}{2}, \frac{5}{3}, \frac{5}{4}, \frac{5}{6}$ and $\frac{6}{5}$).

Problem 4. Find:

a) the greatest common fraction, which corresponds to a domino;

b) the smallest common fraction, which corresponds to a domino;

c) the greatest decimal, which corresponds to a domino;

d) the smallest decimal, which corresponds to a domino.

Answer: a) $\frac{6}{1}$; b) $\frac{0}{1};...;\frac{0}{6}$; c) 6,6; d) 0,0.

Problem 5. How many are the dominoes by which one could represent:

a) a common fraction, equal to $\frac{1}{2}$; b) a common fraction, equal to $\frac{1}{3}$;

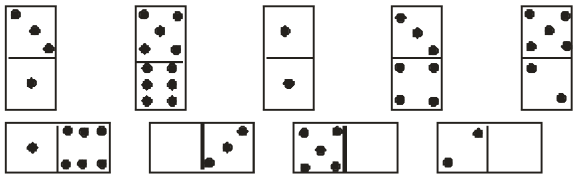
c) a common fraction, equal to $\frac{2}{3}$; d) a decimal, equal to 1,2;

e) a fractional number, equal to $\frac{1}{2}$; f) the digit 0?

Answer: a) three: $\frac{1}{2}=\frac{2}{4}=\frac{3}{6}$; b) two; c) two; d) one; e) four (the representation of the decimal 0,5 should be added to the representation from point a)); f) 7 (six common fractions and one decimal).

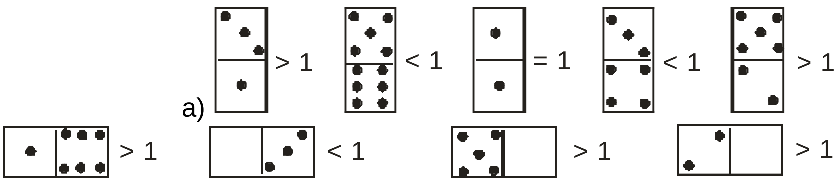
Problem 6. Compare the number 1 with all fractional numbers, which correspond to the dominoes below:

a)
b)



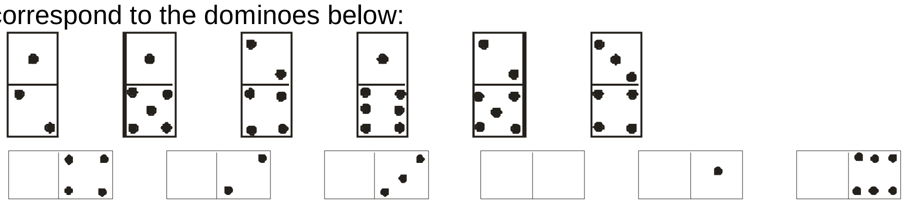
Answer:

a)
b)



Problem 7. Find the difference between the number 1 and all fractional numbers, which correspond to the dominoes below:

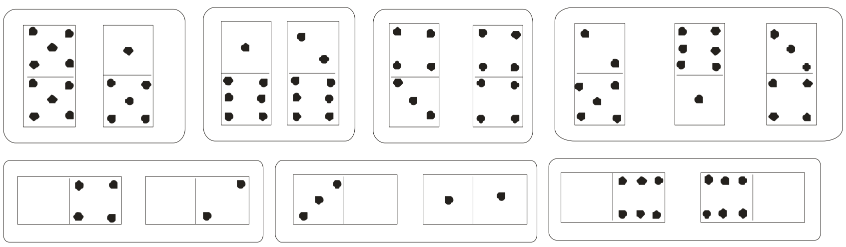
a)
b)

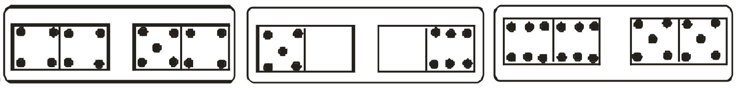


Answer: a) $\frac{1}{2};\frac{4}{5};\frac{2}{4};\frac{5}{6};\frac{3}{5};\frac{1}{4}$; b) 0,6; 0,8; 0,7; 1; 0,9; 0,4.

Problem 8. Find the sum of the fractional numbers, which correspond to the dominoes below:

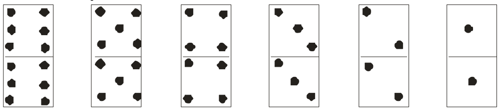
a)
b)

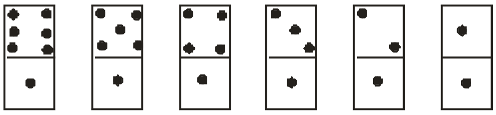


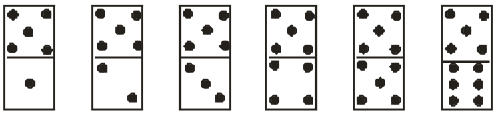
c) 

Answer: a) $\frac{6}{5}; \frac{1}{2}; 2\frac{1}{3}; 7\frac{3}{20}$; b) 0,6; 4,1; 6,6; c) 9,8; 5,6; 12,1.

Problem 9. The sum of the fractional numbers, which correspond to the given dominoes is equal to:

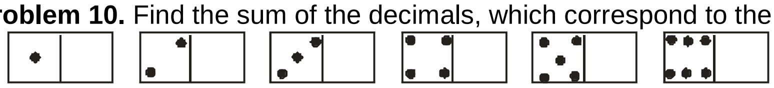
a) 

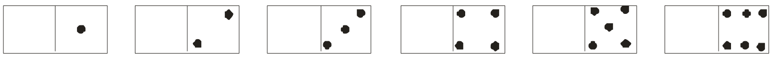
b) 

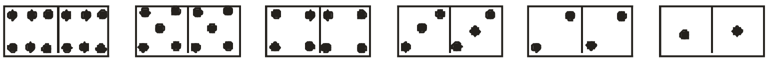
c) 

Answer: a) 6; b) 21; c) $12\frac{1}{4}$.

Problem 10. Find the sum of the decimals, which correspond to the dominoes:

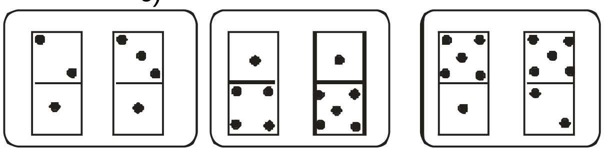
a) 

b) 

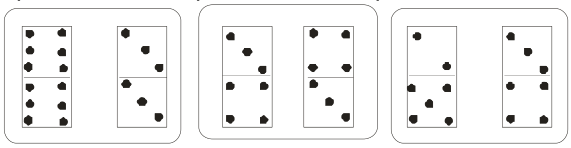
c) 

Answer: a) 21; b) 2,1; c) 23,1.

Problem 11. Compare the common fractions, which correspond to the dominoes:
a) b) c)

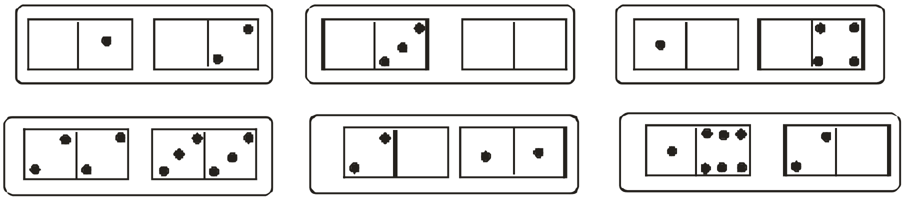


d) e) f)



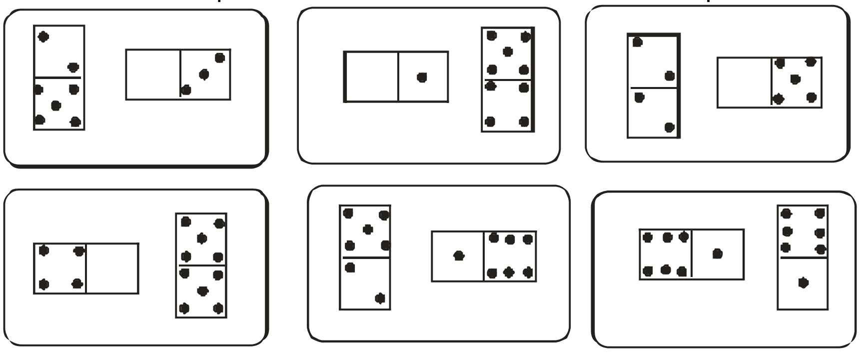
Answer: a) 2 < 3; b) $\frac{1}{4} > \frac{1}{5}$; c) $5 > \frac{5}{2}$; d) 1=1; e) $\frac{3}{4} < \frac{4}{3}$; f) $\frac{2}{5} < \frac{3}{4}$.

Problem 12. Compare the common fractions, which correspond to the dominoes.



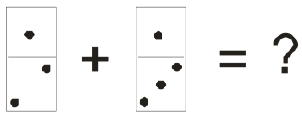
Answer: 0,1 < 0,2; 0,3 > 0; 1 > 0,4; 2,2 < 3,3; 2 > 1,1; 1,6 < 2.

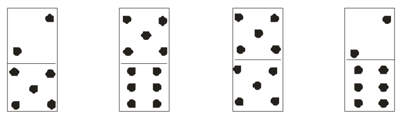
Problem 13. Compare the fractional numbers, which correspond to the dominoes.



Answer: $\frac{2}{5} > 0,3$; $0,1 < \frac{5}{4}$; 1 > 0,5; 4 > 1; $\frac{5}{2} > 1,6$; 6,1 > 6.

Problem 14. By which of the listed dominoes should be replaced the question mark in order to obtain an exact equation?

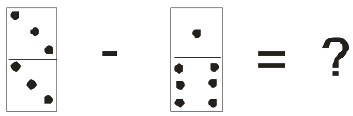


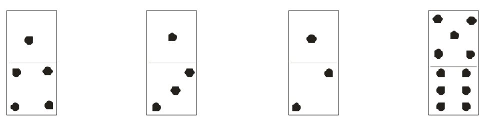


A) B) C) D)

Answer: B).

Problem 15. By which of the listed dominoes should be replaced the question mark in order to obtain an exact equation?

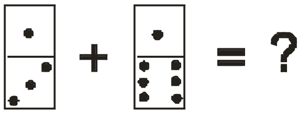




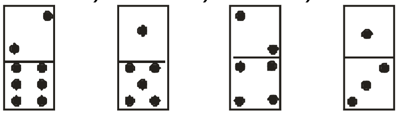
A) B) C) D)

Answer: D).

Problem 16. By which of the listed dominoes should be replaced the question mark in order to obtain an exact equation?



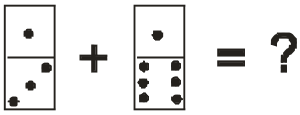
A) B) C) D)



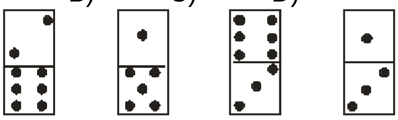
Solution: Add the two common fractions with different denominators $\left(\frac{1}{3}+\frac{1}{6}\right)$ and get $\frac{3}{6}$. Represent the obtained fraction using an irreducible fraction $\left(\frac{3}{6}=\frac{1}{2}\right)$. Then, compare $\frac{1}{2}$ with the common fractions, which correspond to the given dominoes. The conclusion is that $\frac{1}{2}=\frac{2}{4}$.

Answer: C).

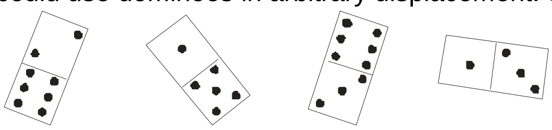
Problem 17. By which of the listed dominoes should be replaced the question mark in order to obtain an exact equation?



A) B) C) D)



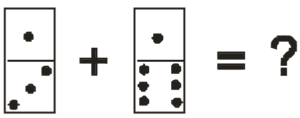
Solution: It is essential in the present problem to discover a domino by which $\frac{1}{2}$ could be represented. Such is the domino in C), but it should be turned up-down. (As an intermediate stage one could use dominoes in arbitrary displacement. See the example.):

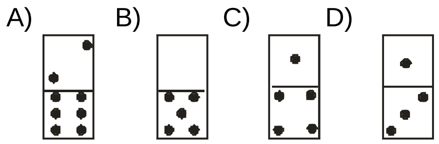


A) B) C) D)

Answer: C)

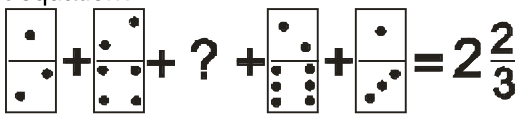
Problem 18. By which of the listed dominoes should be replaced the question mark in order to obtain an exact equation?

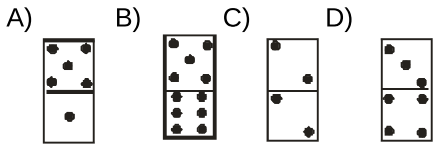




Answer: B) (this domino is used as a model of the decimal 0,5).

Problem 19. By which of the listed dominoes should be replaced the question mark in order to obtain an exact equation?

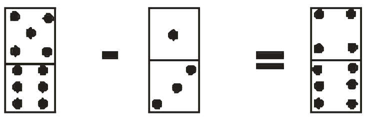




Solution: It is convenient here to require an oral solution. In such a way pupils are directed to look for rational ways of getting the correct answer. In the concrete case one should realize that $\frac{1}{2}+\frac{2}{4}=1$ and $\frac{2}{6}+\frac{1}{3}=\frac{2}{3}$. We get $1\frac{2}{3}$ at the left hand side of the equality and $2\frac{2}{3}$ at the right hand one. Consequently, the question mark should be replaced by a fractional number, which is equal to 1.

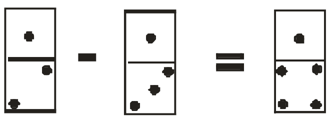
Answer: C).

Problem 20. Replace one of the dominoes by another one in order to obtain an exact equation.



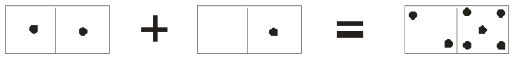
Answer: one of the possibilities is to replace $\frac{5}{6}$ by $\frac{1}{1}$, $\frac{2}{2}$, ... , $\frac{6}{6}$ or by 1,0; a second possibility is to replace $\frac{1}{3}$ by $\frac{1}{6}$; a third possibility is to replace $\frac{4}{6}$ by $\frac{3}{6}$, $\frac{1}{2}$, $\frac{2}{4}$ or by 0,5.

Problem 21. Replace one of the dominoes by another one in order to obtain an exact equation.



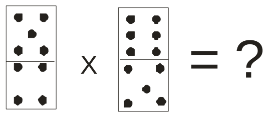
Answer: $\frac{1}{4}$ is replaced by $\frac{1}{6}$ or $\frac{1}{3}$ is replaced by $\frac{1}{4}$.

Problem 22. Replace one of the dominoes by another one in order to obtain an exact equation.

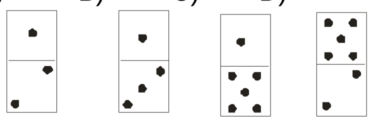


Answer: 2,5 is replaced by 1,2 or by $\frac{6}{5}$; also, it possible to replace 0,1 by 1,4 or 1,1 by 2,4.

Problem 23. By which of the given dominoes should be replaced the question mark in order to obtain an exact equation?



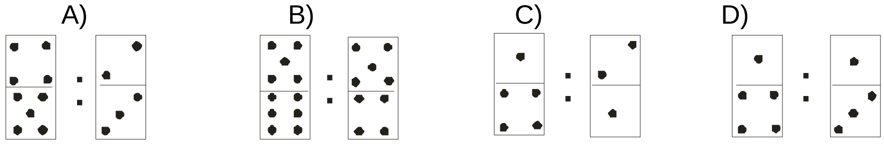
A) B) C) D)



Answer: C) (this domino is used as a model of the decimal 1,5).

Problem 24. Which of the quotients could not be represented by a domino?

A) B) C) D)



Answer: C).

The list of the proposed problems could be extended. The aim of the present paper is to turn the attention to the basic ideas, which are connected with teaching and learning of the topic "Fractional Numbers". As mentioned at the beginning, dominoes could be used in the solutions of other mathematical problems. Some of their various applications will be considered in another paper.

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THE LOGIC IN THE EVOLUTION OF DIDACTIC KNOWLEDGE IN MATHEMATICS EDUCATION

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Abstract. *On the basis of different examples or theoretical generalization an attempt is made to reach a conclusion to separate the stages of knowledge development of Didactics of Mathematics about the ways and tools for their presentation and also for some of their specific properties.*
Key words: *knowledge of Didactic of Mathematics, stages of development*

1. INTRODUCTION

According to the research cited in [2], after the 80s of the previous century until the beginning of the current century four tendencies in education emerged worldwide, one of which is *strengthening education process management by means of educational documentation.*

One growing global tendency from the beginning of the new century (according to a TIMSS study) is to stress the importance of math teachers training.

This training is greatly influenced by the tools for fixing didactic knowledge.

The purpose of the present report is to initiate a systematic and purposeful study of the problem of the evolution of didactic knowledge – to make a distinction between the stages of its development, to define the characteristics and specifications of the different stages, of the tools for fixing didactic knowledge of these stages and the principles for its structuring and organization.

The initial hypothesis is that the presentation of didactic knowledge in historical perspective:

•starts with a *descriptive presentation*, using the natural language;

•the use of *formal models* (of elements of the ratiocinative and predicative calculation or elements of the set theory) follows to present the forms of fixing mathematical knowledge;

•finishes with *content interpretation* of the established formal models described in activities or methods, used and realized by the teacher or the students in the process of mathematics education.

2. THE TOOLS FOR PRESENTING DIDACTIC KNOWLEDGE AT THE THIRD STAGE

In our country by the 60s of the previous century, solutions of didactic problems were presented descriptively by means of the natural language.

At the end of the 60s of the previous century the use of ratiocinative calculation tools in mathematics education was initiated. At first, this tool became an object of independent study [4]. Subsequently, it was introduced in the training of future math teachers (an independent university course was introduced at Sofia University “St. Kliment Ohridski” in 1967 in the curriculum of future math teachers). The purpose of the course is to acquaint the students with this knowledge without using it as a tool when looking into problems of didactics of mathematics education.

Nowadays, when training math teachers, the use of formal models for solving problems of didactics of mathematics education is a fact. When the interpretation is dexterous, the model is transformed into adequate teacher's or students' activities, i.e. ***the model becomes a source of information for the education process management***. The specific content interpretation of the formal models determines the activities of the teacher. The didactic adaptation of the formal models carried out by the teacher in the light of his or her experience, becomes only possible after their realization and then, on this basis, intrinsic methods of action are created. This determines their consistent and purposeful use in identical situations, which externally appear to be seemingly different.

The definitions of mathematical concepts and their theorems are the forms for fixing mathematical knowledge. Namely for this reason, the studies in didactics of mathematics education are directed towards them and their study both by students and for the training of teachers.

A disadvantage of education is the fact that when certain concepts and their theorems are studied, the teacher most often realizes one of the functions – the informational one. The other one – the developing one – which isn't less significant, remains hidden. The realization of the developing function of education requires from the teacher not only to state the facts, but also to point out to the student the ways and the means of acquiring new knowledge on the basis of the acquired knowledge, to point out the situations in which this knowledge could be applied, that is, in general terms – to reveal the creative function of the definitions of concepts and their theorems.

As we said before, these two functions should be distinguished when the teachers are trained, but on a different level. This means that the forms of fixing mathematical knowledge – definition of concepts and theorems – ***initially have to be an object of study***, and after that ***a content interpretation of their formal models should be done in independent teacher's activities or methods***, which he or she has to realize in the teaching and education process.

When future math teachers are trained in the study of the forms for fixing mathematical knowledge, the following models are being used nowadays.

Conclusions are drawn on the basis of schemes $\dfrac{p \to q,\, p}{q}$ (1) and $\dfrac{p \to q,\, \overline{q}}{\overline{p}}$ (2) when the big precondition is a theorem.

Conclusions are drawn on the basis of schemes $\dfrac{p \leftrightarrow q,\, p}{q}$ (3), $\dfrac{p \leftrightarrow q,\, q}{p}$ (4), $\dfrac{p \to q,\, \overline{p}}{\overline{q}}$ (5) and $\dfrac{p \to q,\, \overline{q}}{\overline{p}}$ (6) when the big precondition is a definition or a theorem, providing a necessary and sufficient condition for the existence of the certain concept [10].

Under the specific preconditions, the conclusions are determined from logical, terminological and content point of view.

The small preconditions in the specified schemes (1) - (6) allow for typical situations to be revealed, in which the respective form for fixing knowledge can be used. Image-standard is created for the situation by using typical constructions in Geometry or by considering types of problems which can be solved with this knowledge.

3. EXAMPLES

Example 1. "The principle of complete disjunction" or "Gauber's theorem" are the different names for a logical theorem, which the author of [8] calls a metatheorem for mathematical statements.

Let $p_1 \to q_1$, $p_2 \to q_2$, K, $p_n \to q_n$ are n theorems, such that $v(p_1 \vee p_2 \vee \mathrm{K}\ p_n) = 1$ and $v(\overline{q_i} \vee \overline{q_j}) = 1$ $(v(q_i \wedge q_j) = 0)$ for every $i \neq j$. In that case, $q_1 \to p_1$, $q_2 \to p_2$, K, $q_n \to p_n$ are true statements.

In [1] Gauber's theorem is substantiated by using content preconditions and conclusions. Didactic recommendations of its use in school are presented in the same way.

In [12] the proof of the statement is done completely formally.

Proof. (for ⇐)

Conclusions		*Substantiation*
(!)	q_1	*Given*
(2)	$\overline{q_1} \vee \overline{q_2}$	*Given*
(3)	$\overline{q_2}$	*Rule for elimination of disjunction* $\frac{p \vee q, \overline{p}}{q}$ *u* (1)
(4)	$p_2 \to q_2$	*Given*
(5)	$\overline{p_2}$	*Modus tollens and* (4)
(6)	$\overline{q_1} \vee \overline{q_3}$	*Given*
(7)	$\overline{q_3}$	*Rule for elimination of disjunction* $\frac{p \vee q, \overline{p}}{q}$ *and* (6)
(8)	$p_3 \to q_3$	*Given*
(9)	$\overline{p_3}$	*Modus tollens and* (7)
(10)	$p_1 \vee p_2 \vee p_3$	*Given*
(11)	$p_1 \vee p_2$	*Rule for elimination of disjunction* $\frac{p \vee q, \overline{p}}{q}$ *and* (9)
(12)	p_1	*Rule for elimination of disjunction* $\frac{p \vee q, \overline{p}}{q}$ *and* (5)

The proposed verbal and symbolical proofs can be used for the training of math teachers.

In order for the students to be able to apprehend the explanations, which are concrete manifestations of the verbal generalized substantiation of the statement given in [1], the presence of knowledge is required since it provides the small precondition when the argument is modus ponens, as the big precondition is the statement

$$(p_1 \vee p_2 \vee \ldots \vee p_n \wedge \overline{q_i} \wedge \overline{q_j} \wedge i \neq j \wedge$$
$$p_1 \to q_1 \wedge p_2 \to q_2 \wedge \ldots \wedge p_n \to q_n) \to \qquad (\oplus\ \oplus).$$
$$q_1 \to p_1 \wedge q_2 \to p_2 \wedge \ldots \wedge q_n \to p_n$$

The small preconditions are the statements (1), (2) and (3).

$$p_1 \vee p_2 \vee \ldots \vee p_n \qquad (1)$$
$$\overline{q_i} \wedge \overline{q_j},\ i \neq j \qquad (2)$$
$$p_1 \to q_1, p_2 \to q_2, \ldots, p_n \to q_n \qquad (3)$$
$$q_1 \to p_1, q_2 \to p_2, \ldots, q_n \to p_n \qquad (4)$$
$$\frac{(1) \wedge (2) \wedge (3) \to (4),\quad (1) \wedge (2) \wedge (3)}{(4)}$$

Since in this case statements (3) are proved, in order to draw conclusions (4), purposeful exercises on specific level are required, so that preconditions (1) and (2) can be acquired, and they are connected with the formation of skills for breaking the set into subsets and of skills for determining truthfulness of the argument with the given logical structure. After these skills have been formed, by means of an algorithmic prescript it is postulated that under these conditions all statements contrary to the proven ones are true. This algorithmic prescript is for the students a tool for:

- summary of their knowledge;
- systematization of their knowledge.

Example 2. Let a, b, c are sides of ΔABC, and R is the radius of the circumcircle about the triangle.

- If $a^2 + b^2 + c^2 = 8R^2$, then the triangle is right-angled.
- If $a^2 + b^2 + c^2 > 8R^2$, then the triangle is acute-angled.
- If $a^2 + b^2 + c^2 < 8R^2$, then the triangle is obtuse-angled.

The truthfulness of these statements - sufficient conditions follows from the basis of the logical statement in hand and from the truthfulness of the following theorems – necessary conditions.

- ✓ If ΔABC is right-angled, then $a^2 + b^2 + c^2 = 8R^2$.
- ✓ If ΔABC is acute-angled, then $a^2 + b^2 + c^2 > 8R^2$.
- ✓ If ΔABC is obtuse-angled, then $a^2 + b^2 + c^2 < 8R^2$.

Example 3. If for a triangle ΔABC we are given that $OM = \frac{1}{3}R$, where points M and O are respectively the triangle centroid and the center of a circumcircle with a radius R, then the triangle is right-angled.

To prove this statement we use the argument that the power of point M with respect to the circumcircle is $p(M) = -\frac{1}{9}(a^2 + b^2 + c^2)$ and the first of the sufficient conditions from the previous example. The presence of the three sufficient conditions implies that the problem can be summarized as follows.

For a triangle ΔABC we are given that $OM \rho \frac{1}{3}R$, where points M and O are respectively the triangle centroid and the center of a circumcircle with a radius R, and ρ

is the relation “=”, “<” or “>”. Then ΔABC is respectively right-angled, acute-angled and obtuse-angled [9].

CONCLUSION

The achievement of the objectives of the present report requires many years of creative work most probably by a group of people.

Of course, within the range of such a study, we cannot eliminate the idiosyncratic acquisition and analysis of this knowledge, but looking at it from a temporal distance we can draw some conclusions against the background of the general development of the pieces of didactic knowledge and their relative connections.

The initial conclusions given so far show that, indeed, in our country the training of teachers is not put on an empirical basis, but is based on theoretical facts and generalizations. What is emerging is a tendency to strengthen the management of the process of teachers training on the basis of the already created theory for mathematics education.

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THE MATHEMATICAL MODELING – AN INPORTANT ASPECT IN UNIVERSITY TRAINING OF PHYSICS MAJORS

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***Abstract**: The paper deals with and analyzes the specific peculiarities of ordinary differential equations as mathematical models in the process of acquiring knowledge in physics. Special attention is paid to the both parts of mathematical modeling, seen as an explanation: the construction of differential equation and revealing its physical meaning and content. This problem is analyzed in close connection with the problem of the limits and conditions for applying these mathematical models. The acquisition and implantation of this method by students in their own cognitive activities is discussed as a very important didactic task in education.*

***Keywords:** mathematical modeling, ordinary differential equations, physics majors*

1. INTRODUCTION

Development of science is unimaginable without using such a universal cognitive method like the method of modeling. Its important features consist in the fact that it is applicable only in close logical relation with other methods of scientific cognition, it is used by all sciences and at all stages of research. By using models and analysis of models, the more complex and abstract things are reduced to simpler and more available things, the invisible is reduced to the visible, the unknown is reduced to the known.

Model – this is "an imaginary represented or materially implemented system which, by reflecting or reproducing the object of study, is capable to substitute it such that study of the system gives us new information about this object" [1].

Modern physics, like all other natural sciences, is characterized by the process of increasingly profound "mathematization", which is done at both empirical and theoretical cognitive level. This mathematization is expressed in the fact that these sciences cannot be developed and research in these sciences cannot be done without using the formal language of mathematics, without applying various mathematical concepts, ideas and methods, that is, mathematics is both the language and logic of natural sciences. Thus, the qualitative description of objects and phenomena is complemented by a quantitative description. It provides the opportunity to discover new, unknown sides of objects and phenomena, to predict the course of the going processes, and this is a prerequisite that they be managed and successfully applied in practice in favor of humans.

When forming knowledge in physics – both as a science and as a school course – an important role is played by this aspect of mathematization, which is connected with the construction of various mathematical models of real and abstract situations that are studied. Mathematical modeling is an inseparable part of the research process and it takes leading place in the general scheme of the cognitive process. Its heuristic role and logical value are incomparable greater than those of other methods of scientific cognition. This discussion shows that the method of mathematical modeling should find its adequate reflection in education of physics majors.

One of the basic aims of the Mathematical methods of physics (MMP) course is to reveal the role and importance of ordinary differential equations (ODEs) as models in the formation of scientific knowledge. This necessitates application of such educational methods which allows students to acquire properties of the ODEs and methods of their integration not as an end in itself but as an important part as well as in the context of such a universal research approach like the method of mathematical modeling. Thus, knowledge and skills which are formed during the educational process, assume high methodological value and they are powerful factor for the formation of positive motivation to study. Moreover, they correspond to another important trend in modern education: it should be orientated not only to acquiring classified and systematized information and knowledge by students but mainly to acquiring cognitive methods and approaches [2].

The following important question arises: Which are basic aspects of this type of mathematical modeling, that lecturer has to reveal in front of students during the educational process?

2. SPECIFIC FEATURES OF ORDINARY DIFFERENTIAL EQUATIONS AS MODELS OF FORMATION OF PHYSICS KNOWLEDGE

In the history of physics, three basic types of mathematical models can be found – geometric, dynamic and statistical models. Dynamic models, in fact, rely on the apparatus of differential equations. Their basic purpose is expressed in the fact that they allow to do single-valued predictions, which means to obtain completely determined, unique law for the considered phenomenon. This law is determined only by the differential equation and given specific additional conditions. Therefore this type of mathematical models gives the concept of continuous deterministic process. On the other hand, ODEs are sign-symbolic models, that is, they "are records, reflection of the structure and characteristics of the object under modeling by means of signs-symbols of some artificial (formal) language" [3] – in this case, the language of mathematics. Hence, ODEs are perfect models of various problem situations and in this sense they are materialized substitutes of real objects, phenomena and processes. From psychological-pedagogical point of view, it is very important for students to understand that these models are mathematical structures, in which the real and concrete relations are replaced by abstract mathematical relations, that is, each differential equation is a high level abstraction, which reflects characteristics of the real world more adequate or less adequate, and it is useful for the cognition to such an extent, as far as it reproduces the reality correctly, that is, essential features of the considered objects, phenomena and processes. In this sense, mathematical models of this type can also be considered as a way for demonstrativeness in the area of theoretical thinking. In this case, we mean the so called abstract demonstrativeness. It allows knowledge to be presented at the level of the essence or at the level of the law-governed connections and relations of what cannot be perceived by our senses, in general, of the latent and inaccessible course of processes and phenomena. Mathematical equations, including ODEs, which are quantitative models, are high level form of abstract demonstrativeness. They have high information and operational content, have high level of explanatory value and they cannot be understood and successfully applied in practice without specific training. First of all, this training should be directed to acquiring the conventionality of each separate symbol, that is, what exactly is written by this symbol, what is its meaning, to which rules it is subordinated and by which algorithms it is processed. After that, the connection between separate symbols in the differential equation should be discovered and understood, as well as the conformity to a law brought by it, the rules should be acquired by which the particular differential equation is operated on, and only then to go to problem solving. Therefore, operating with symbol models in the cognitive process presumes a complex system of intellectual activities. Formation of such skills requires psychologically satisfactory educational methods by which the abstract demonstrativeness is transformed to an object of cognition. Introduction of abstract models and finding appropriate approaches for revealing their semantics is one of the ways for acquiring intellectual activities, which change the quality of psychic processes and allow the transition to higher degree of thinking – the level of abstract and generalized thinking

3. STRUCTURE OF THE MATHEMATICAL MODELING AND ITS PRACTICAL IMPLEMENTATION IN EDUCATION

The fact that ODEs are powerful tool of knowledge requires students to be familiar with both parts of mathematical modeling in the educational process: mathematical scheme, that is, one or several mathematical formulas, in the particular case ODEs, and interpretation of this scheme, that is, rules of its interrelation with the particular physical situation.

First part is connected with the construction of the model, that is, composition of the ODE, when it turns out to be a product of knowledge. This aspect of mathematical modeling can be acquired in practice if during the educational process appropriate conditions are provided for active inclusion of students in joint activities for solving physics problems. The purpose is: on the basis of a small number of precisely selected problems to implement such problem situations which reveal the origin of knowledge, show their genesis and logic of their development. As it is known, problem situations arise in the process of activity of the subject, which is directed to a particular object, when subject encounters difficulty or obstacle. The substance of each problem situation consists in the fact that "it is not simply a difficulty, obstacle in the activity of the subject but it is realized by the subject difficulty, the way of removing of which he wants to find out" [4]. This fact is an important premise for the origination of intellectual activity directed to detailed analysis of the situation, finding out its components, connections and relations among them, the nature and special features of the obstacle. When solving physics problems, the detailed analysis allows modeling of the considered physical situation and its description using the formal language of mathematics, that is, transformation of the situation to a particular mathematical problem. This approach allows the students to trace the origin of differential equations during the research process, to reveal the mechanism of their construction, to understand the meaning and the necessity of learning this mathematical tool.

Second part of mathematical modeling is not less important. It has clearly expressed substance: to obtain particular corollaries of the mathematical model. For practical implementation of this aspect, it is not enough that students have acquired profoundly methods for solving various types of ODEs. Interpretation of both differential equation and its solutions is of crucial significance. It allows to establish limits and conditions of applicability of the mathematical model, to reveal its physical meaning, its physical substance. That is why a required element of the educational methods should be precise geometric interpretations of various types of solutions which are obtained when integrating the differential equation. Their use as a visual tool helps decoding of the information which is contained in the sign-symbolic form of representation of these abstract scientific concepts, allows to connect them with the actual reality, and therefore it helps for their correct understanding. Graphic mappings are demonstrative tools of artificial type. They do not reflect visual features of the object under study but they are directed to conditional representation and fixing of the abstract, therefore they are variety of the abstract demonstrativeness. Their cognitive value is also strengthened by the fact that they are not connected uniquely with a single object but they unite in itself the conformity to a law in a group or class of objects. The interpreting function of geometric models can be implemented in the educational process only on condition that they are thoroughly analyzed, during which students clarify and understand the conventionality of the signs, symbols, draft or mapping of the method, by which a demonstrative construction can be made. In order to obtain qualitative geometric representations, various specialized computer programs can be used such as MATHEMATICA, ORIGIN, MATHCAD, etc.

It is important to emphasize in the educational process that when implementing the second part of the considered type of mathematical modeling, differential equation plays the role of tool of knowledge. In this case, of significant meaning is to establish the adequacy (isomorphism) of the constructed mathematical model to the physical situation given in the problem. That is why the final stage in the cognitive activity of students should be directed to the investigation of theoretical conclusions of the model. This presumes a profound analysis, during which the information about obtained solutions of the differential equation is compared with the text of the problem, that is, the transfer of knowledge from the model to the object is accomplished in order to find the solution we are looking for.

Therefore, in order to acquire basic structural elements of the method of mathematical modeling, students should take part in the specific activity in which they perceive and understand ODEs as sign-symbolic models of real and abstract problem situations, form skills to construct differential equations, use expediently selected ways for integrating them and interpret the obtained solutions. An important additional condition is the selection of problems with applied nature which are basis for acquiring mathematical modeling. These problems have to reveal other important aspects of mathematical modeling connected with the following possibilities: the same differential equation is model of various situations; conclusions of the mathematical model are obtained by using various methods for solving the respective differential equation; various models are constructed for studying the same situation.

From methodological point of view, the third case is very important. It has two versions of realization, which students are required to analyze:

- differential equations are of various type but various models lead to the same result, that is, with the same degree of accuracy differential equations reflect substance of the situation, given by the text of the problem [5];
- differential equations are of various type and various models with different degree of accuracy reflect substance of the situation, given by the text of the problem [6]. This corresponds to the frequently met in the science case when a model cannot describe all observed phenomena, properties and characteristics of the object that is why necessity of improvement or exchange of the model arises.

4. CONCLUSIONS AND INFERENCES

Practice shows that acquisition of mathematical modeling should become a required element of the curricula of many courses, studied by physics majors. Presently, the emphasis in curricula is on search and practical implementation of the optimum balance between various systems of concepts and the methods of their acquisition, and the problem of acquiring ways of how to think, how to acquire new knowledge, how to create, that is, how to acquire knowledge-methods, is less emphasized. However, it is indisputable that the social and individual meaning of acquisition of knowledge consists in the ability to apply them in practice, to solve various theoretical and practical problems through them, that is, people use them for the change of life and the world. In the presence of strengthened processes of abstractness, differentiation and systematization of the scientific knowledge, this social requirement can be satisfied only on condition that orientation towards acquisition of methods of thinking and practical action becomes dominating in the curricula. In this sense, mathematical modeling, as a universal cognitive method, should find its adequate reflection in physics majors education. This fact poses another important problem. It refers to the methods, ways and means by which knowledge-methods should be acquired in the educational process. A possible version is method to be acquired as instructions, as a prepared model for organization of the intellectual activity, demonstrated by the lecturer. The second possible version is embraced in the proposed in this paper methodology of education. It has the specific feature that acquisition of the method-knowledge is not carried over from outside, is not taken without any effort but it is given by posing a problem, which programs the operational content of the cognitive activity. This approach is based on the self-sufficient and active acquisition of personal experience by students. This leads to specific logical-operational structure of the intellectual activity and it is a prerequisite for forming the necessary qualities of the psychical processes and properties. Therefore, from psychological point of view, this approach allows to realize in practice the developing function of education.

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ANALYSIS OF RESULTS OF EXPERIMENTAL COMPUTER-AIDED PHYSICS TEACHING

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***Abstract**: The report contains analysis and interpretation of results of pedagogical experiment, using computer-aided educational technologies in physics teaching in secondary school education, chapter ”Mechanical oscillation and waves“. The mathematical and statistical methods of processing the experimental data are described and reasoned. Also, basic conclusions of the analysis of the pedagogical research are formulated.*

***Keywords:** methodology of teaching physics, computer-aided educational technologies, pedagogical experiment, secondary school*

At times of continuous and intensive development of computer technologies is created real premises for the creation of new technology to organize the educational process and exchange of information in it.

The purpose of the conducted pedagogical experiment [1] is to determine the effectiveness of secondary school physics teaching using computer-aided educational technologies.

The pedagogical experiment was conducted in NHS ”St. St. Cyril and Methodius“ – Blagoevgrad in the period 2001/2002–2004/2005 school years with students of 9th and 10th class that studied the chapter ”Mechanical oscillations and waves“ following the curriculum of 9th class, first level [3]. The extract of the experiment was formed from 417 students, divided in two groups – control and experimental. In control group educational process was conducted using traditional methodology of physics teaching. The students of experimental group were taught by appling the created methodological conception for complex and purposeful use of computer-aided educational technologies of secondary school physics teaching, in the base of which is the created methodologically founded system of didactic software products that can be applied when teaching chapter ”Mechanical oscillations and waves“ in 9th class, first level.

Processing and analysis of the experimental results was done using the following mathematical and statistical methods:

- descriptive statistics – determination of arithmetic mean, standard deviation, standard error, frequency tables, histograms, diagrams, etc.;
- hypotheses verification methods about the type of distribution – Pearson’s chi-square test and Kolmogorov–Smirnov test;
- hypotheses verification methods about the statistical importance of differences in the arithmetic means of statistical characteristics – parametrical (Student’s *t*-test, Lavene’s test, *F*-test), non-parametrical (Mann–Whitney test and Wilcoxon test), dispersion analysis (Kruskal–Wallis test) and others [2].

To verify of the hypothesis of the pedagogical research a null statistical hypothesis H_0 is formulated: there are no statistically significant differences between the educational achievements of the students of the experimental and control groups, the different aspects that were evaluated through various forms of control and assessment (oral examination, solving of problems, preparation of report under a given topic, conducted laboratory exercises, concluding test, re-test).

$$H_0:\ \mu = \mu_0$$

An alternative statistical hypothesis H_1: there are statistically significant differences between the educational achievements of the students of the experimental and control groups, the different aspects that were evaluated through various forms of control and assessment (oral examination, solving of problems, preparation of report under a given topic, conducted laboratory exercises, concluding test, re-test).

$$H_1:\ \mu \neq \mu_0$$

To test the statistical hypothesis a variety of different statistic sub-hypotheses reflecting different aspects of the pedagogical experiment were formulated, using mathematical and statistical methods to process and analyze experimental results (SPSS 11.0).

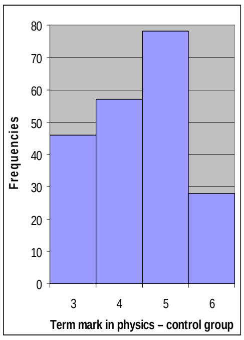

Fig. 1.

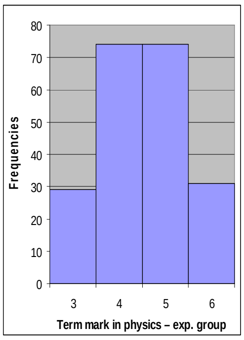

Fig. 2.

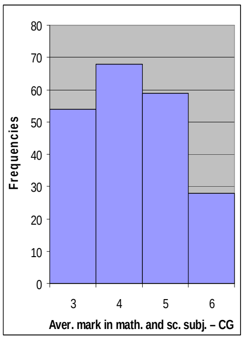

Fig. 3.

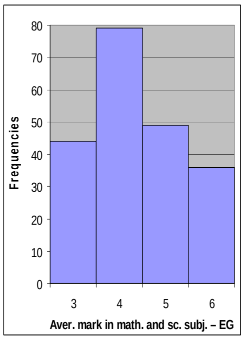

Fig. 4.

For the evaluation of the homogeneity of control and experimental groups in reference of entry level, two statistical characteristics were formulated: X – represents the term mark in physics, and Y – average mark in mathematics and science subjects (chemistry and biology). The histograms of X and Y characterizing the control and experimental groups are presented in fig. 1–4.

The Pearson's chi-square test shows that the distribution of these characteristics is not normal. The empiric characteristics of X and Y for each of the groups are presented in table 1.

Table 1.

$\chi^2_{emp.}$	Experimental group	Control group
X	7,168	8,530
Y	11,701	5,831

$$\chi^2_{0,05}(1) = 3{,}84 \qquad \chi^2_{emp.} > \chi^2_{theor.}$$

Along with the study of distribution type of characteristics X and Y with the aim to determine the homogeneity of groups in regards to entry level, basic numeral characteristics were calculated – arithmetic mean, median, mode and standard deviation for both statistical characteristics (table 2). Numeral characteristics in regards to measurement of the central trend, as well as dispersion measurement do not differ too much.

Table 2.

Numeral characteristics				
	X – EG	Y – EG	X – CG	Y – CG
Number of observations	208	208	209	209
Arithmetic mean	4,510	4,370	4,420	4,292
Median	5,000	4,000	5,000	4,000
Mode	4,000	4,000	4,000	4,000
Standard deviation	0,911	1,004	0,978	0,998

Table 3.

	Term mark in physics	Average mark in mathem. and science subjects
Mann–Whitney (U)	20739,000	20888,500
Significance level (p)	0,396	0,473

For statistical importance verification of the differences in the arithmetic means of *X* and *Y* characteristics a non-parametric criterion was used, because the empiric distributions of control and experimental groups do not belong to a normal class. The results of Mann–Whitney test show that at error probability of first kind $\alpha = 0{,}05$, there is no statistically significant difference between the arithmetic means of control and experimental groups for both characteristics (table 3). The solution is based on values of significance level *p* that are higher than error probability $\alpha = 0{,}05$ for both statistical characteristics. This proves that in regards to the studied characteristics, both groups are homogeneous.

Taking in view that students from different classes take part in the pedagogical experiment, a check of homogeneity of control and experimental groups in different years was done. A monofactor dispersion analysis was used, where factor variable were years, and resulting variables are *X* and *Y* characteristics. The results of the dispersion analysis non-parametric Kruskal-Wallis test (tables 4 and 5), applied to the empiric data, that were not normally distributed, show that differences between the arithmetic means of *X* and of *Y* characteristics are not statistically significant ($p > \alpha = 0{,}05$). This proves the homogeneity of the studied groups in regards to entry level, no matter in which year the pedagogical experiment was undertaken.

Table 4.

Results of the dispersion analysis			
	School year	Number of observations	Mean rank
Term mark in physics	2002-2003	139	197,13
	2003-2004	136	218,57
	2004-2005	142	214,37
	Total	417	
Average mark in mathem. and science subjects	2002-2003	139	198,45
	2003-2004	136	218,95
	2004-2005	142	212,84
	Total	417	

Table 5.

Non-parametric Kruskal-Wallis test of dispersion analysis		
	Term mark in physics	Average mark in mathem. and science subjects
χ^2	2,707	2,224
df	2	2
Significance level (p)	0,258	0,329

The results of the pedagogical experiment are reported through various forms of control and assessment of educational achievements of students (oral examination, solving of problems, preparation of report under a given topic, conducted laboratory exercises, concluding test), the basic descriptive characteristics of which are presented in table 6.

Table 6.

Numeral characteristics						
Groups		Oral examination – mark	Solving of problems – mark	Preparation of a report – mark	Conducted laboratory exercises – mark	Concluding test – mark
Control group	Number of observations	209	209	209	209	209
	Arithmetic mean	4,02	3,97	5,25	4,01	4,10
	Standard deviation	1,166	1,078	0,788	1,054	1,043
	Standard error	0,081	0,075	0,054	0,073	0,072
Experimental group	Number of observations	208	208	208	208	208
	Arithmetic mean	4,77	4,73	5,74	4,77	4,91
	Standard deviation	0,928	0,956	0,484	0,897	0,880
	Standard error	0,064	0,066	0,034	0,062	0,061

The results show higher average grades in all forms of control and assessment of student educational achievements from experimental group. To verify the hypotheses about empiric distributions type the Kolmogorov-Smirnov test was used, as well as the Pearson's chi-square test. The results of their application do not give ground to accept the suppositions of empiric distribution normality.

Table 7.

	Oral examination – mark	Solving of problems – mark	Preparation of a report – mark	Conducted laboratory exercises – mark	Concluding test – mark
Mann–Whitney (U)	13886,50	13332,00	14494,00	13149,00	12207,50
Significance level (p)	0,0002	0,0001	0,001	0,0001	0,0004

To verify the hypotheses about statistic significance of arithmetic means differences in the statistical characteristics it is necessary to use the non-parametrical Mann–Whitney test and these results are shown in table 7. They allow the hypothesis about the availability of statistically significant differences between the average student achievements of control and experimental groups in relation to different forms of control and assessment to be accepted as not contradicting to empirical data ($p < \alpha = 0{,}05$).

Additional conclusions in regards to the results of the suggested method are reached by comparing the distribution of student educational achievements in various forms of control and assessment. These results are graphically shown in fig. 5–9, relevant to the characteristics of the control and experimental groups. They are a basis to conclude that the statistically higher achievements of the experimental group are result of the improvement of the achievements of all students, not only of poor or good ones.

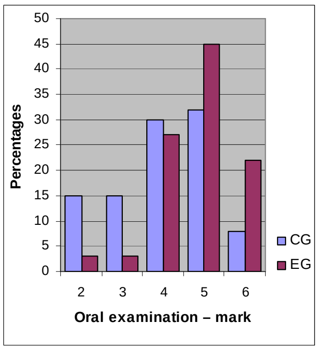

Fig. 5.

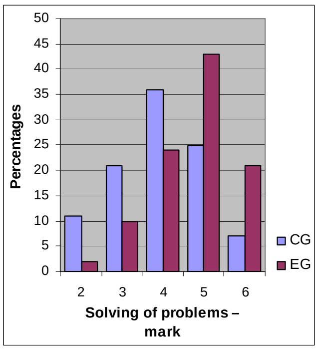

Fig. 6.

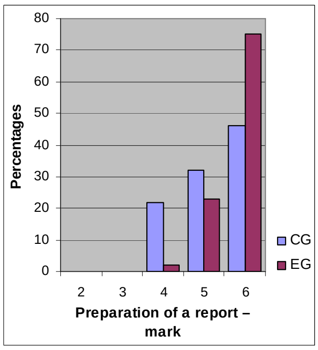

Fig. 7.

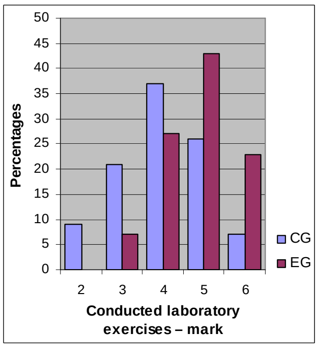

Fig. 8.

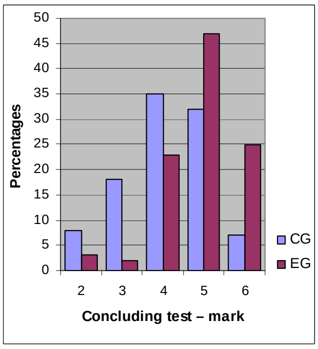

Fig. 9.

When analyzing the results of the test as number of points, the data is grouped according to Sturgest formula and the concrete test specifics. The histograms about control and experimental groups were built at an interval width of *h*=4 points (fig. 10 and 11). The comparison of distributions shows that there is an aberration of empiric distribution center towards higher values for the experimental group.

The verification of the hypotheses about type of distributions shows that there is no ground to reject the normality suppositions ($p > \alpha = 0{,}05$). The results of the Kolmogorov–Smirnov test are presented in table 8.

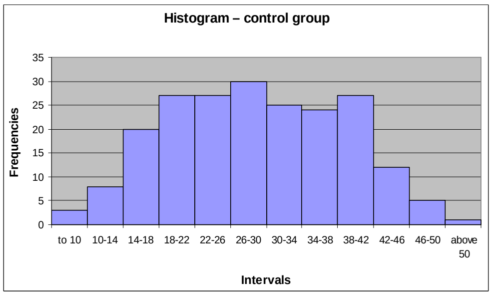

Fig. 10.

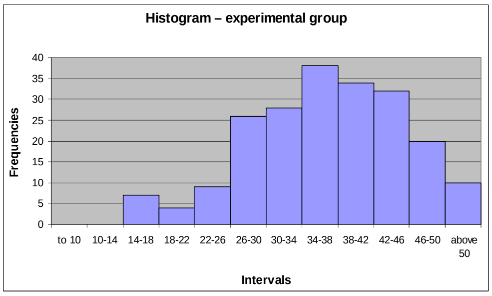

Fig. 11.

Table 8.

Kolmogorov-Smirnov test			
		Control group	Experimental group
Numeral characteristics	Number of observations	209	208
	Arithmetic mean	29,3589	37,0769
	Standard deviation	9,6327	8,8456
Maximum differences	Absolute	0,057	0,051
	Positive	0,057	0,031
	Negative	– 0,057	– 0,051
Z_{emp}.		0,829	0,742
Significance level (p)		0,498	0,640

In that case to verify the hypothesis for parity of the arithmetic means the parametric Student's *t*-test was applied, and it is preceded by a verification of a hypothesis about the parity of dispersions (Lavene's test) (tables 9 and 10).

Table 9.

Numeral characteristics		
Groups		
Control group	Number of observations	209
	Arithmetic mean	29,36
	Standard deviation	9,633
	Standard error	0,666
Experimental group	Number of observations	208
	Arithmetic mean	37,08
	Standard deviation	8,846
	Standard error	0,613

Table 10.

		Lavene's test		Student's t-test		
		F	Significance level (p)	t	df	Significance level (p)
Number of points from test	Equal dispersions	4,498	0,035	– 8,521	415	0,0002
	Different dispersions			– 8,522	412,345	0,0002

Because the significance level of the Lavene test is 0,035, i.e. less than the error probability of first kind $\alpha = 0{,}05$, the difference in the two groups dispersions is statistically significant. This shows existence of statistically significant dispersion in both groups and requires the use of Student *t*-test in suggestion of different dispersions. In that case the results of hypothesis verification about the parity of the arithmetic means are presented in the lower row of table 10 (for different dispersions) and shows statistically significant differences in the average number of points ($p = 0{,}0002 < \alpha = 0{,}05$).

The analysis of the test results as number of points shows the availability of statistically significant differences between the educational achievements of the students from the two groups – control and experimental one.

To determine student knowledge and skills durability and to follow up the stability in the results from the testing after the experimental education another follow up re-testing was performed four weeks after the experiment. The histograms of control and experimental groups are presented in fig. 12 and 13.

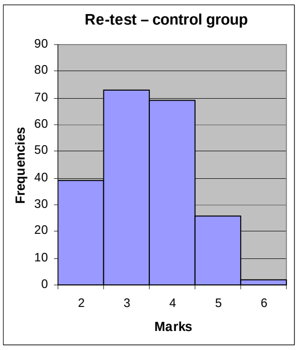

Fig. 12.

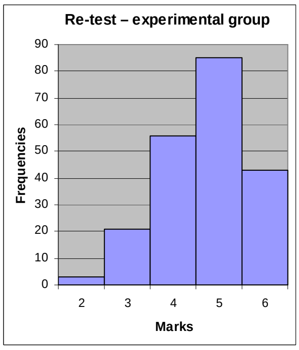

Fig. 13.

The descriptive statistics, presented in table 11, shows that the arithmetic mean of the experimental group is higher, which is a proof of retain of the results of the experiment in regards with the acquired knowledge and skills. The standard deviation, as well as the results of the testing, have lower value in the experimental group.

Table 11.

Numeral characteristics		
	Re-test – control group	Re-test – experimental group
Number of observations	209	208
Arithmetic mean	3,42	4,69
Standard deviation	0,96314	0,95883
Standard error	0,06662	0,06648

Table 12.

	Re-test
Mann–Whitney (U)	8190,00
Significance level (p)	0,0003

The application of the non-parametrical Mann–Whitney test shows that there is statistically significant difference between the average educational achievements of the control and experimental groups $(p = 0{,}0003 < \alpha = 0{,}05)$, which is a durability exponent of the of knowledge and skills acquired through the applied methodological system in physics teaching (table 12).

Despite of retaining the statistic significance of the differences in student achievements in both groups during the re-testing, there is still a reduction of achievements in comparison with results from original test. This reduction is based on the timing between both tests, and it is important to define whether these reductions are statistically significant. Because these results belong to same students, to verify the statistic significance of the difference in the arithmetic means the non-parametrical Wilcoxon test of dependent extracts was used (tables 13 and 14).

Table 13.

	Re-test control group – Test control group
Z	$-2{,}815$
Significance level (p)	0,005

Table 14.

	Re-test experimental group – Test experimental group
Z	2,120
Significance level (p)	0,240

The results of Wilcoxon test show the statistically significant reduction in student achievements in the control group $(p = 0{,}005 < \alpha = 0{,}05)$ (table 13). The reduction of student achievements in the experimental group in the re-testing is not statistically significant $(p = 0{,}240 > \alpha = 0{,}05)$ (table 14).

The mathematical and statistical processing of the experimental results gives ground to reject the null statistical hypothesis and to accept its alternative, that there are statistically significant differences in the educational achievements of students in control and experimental groups. This proves the pedagogical research hypothesis.

The use of a concrete model of mathematical and statistical processing of the results from the conducted pedagogical experiment and its qualitative analysis prove that the application of the developed methodological system (conception and didactic software products) for complex and purposeful use of computer-aided educational technologies increases the effectiveness of secondary school physics teaching.

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COMPARATIVE EFFECT OF NUCLEOSIDE ANALOGUES AGAINST REPLICATION OF HERPES SIMPLEX VIRUS TYPE 1 IN VITRO

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1. INTRODUCTION

Acyclovir, 9-[(2-hydroxyethoxy)methyl] guanine (ACV) is an acyclic guanine nucleoside analogue that is widely used clinically as an antiherpetic agent. Its limited absorption (15%-20%) in humans after oral administration prompted the search for prodrugs possessing higher bioavailability [1-2]. A possible way to increase the bioavailability is to modify the known antiviral drugs with various amino acids. The L-valyl ester of aciclovir (valaciclovir) with bioavailability of 60% is obtained in this manner [3-4].

Modification of anti-herpes agents like aciclovir by peptidomimetics, whose chemical structure is different from the natural peptides but have the same ability to interact with specific receptors, is of definite interest [5].

Modification of acyclovir with thiazole containing dipeptide mimetics derived from alanine and leucine can enhance the oral bioavailability of the parent drugs.

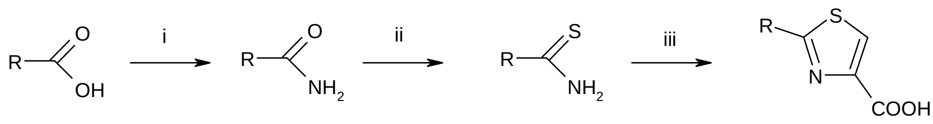
Moreover, such prodrugs may have longer half-live time, for the identification of the peptidomimetics (wich are pro-moiety) by the cellular enzymes may be hindered. The slower release of the prodrug could result in reduced toxicity.

The aim of this study was to design and to synthesize of new thiazole containing amino acids esters of acyclovir and to explore their activity on the HSV-1.

2. MATERIALS AND METHOD

2.1. Synthesis of thiazole containing amino acids

Boc-2-Ala-thiazole-4-carboxylic acid and Boc-2-Leu-thiazole-4-carboxylic acid was synthesized according to **Figure 1** [6-7]. The amide **2** was obtained following Pozdnev's method [8] from Boc-Ala-OH and Boc-Leu-OH (**1**) and converted into thioamide **3** by Lawesson's reagent [9]. Cyclocondensation of **3** with 3-bromo-oxo-propionic acid [10] leads to **4.**



R=Boc-Ala-, Boc-Leu-
(**i**) $(Boc)_2O$, NH_3HCO_3; (**ii**) Lawesson`s reagent; (**iii**) 3-bromo-2-oxo-propionic acid

Fig. 1. Synthesis of Boc-Ala-thiazole-4-carboxylic acid and Boc-Leu-thiazole-4-carboxylic acid

2.2. Synthesis of amino acids of acyclovir

Synthesis of thiazole containing amino acid esters of acyclovir involved (**5**) formation of N-Boc protected amino acid anhydrides, (**6**) coupling of the N-Boc protected amino acid anhydride with acyclovir, and (**7**) deprotecting the amino group of the amino acid ester of acyclovir.

A mixture of N-Boc protected amino acid containing thiazole ring (Boc-AA) and dicyclohexyl carbodiimide (DCC) in dimethylformamide (DMF) with ratio 1:2 DCC/Boc-AA was stirred for 1 hr at 0℃ under nitrogen atmosphe re. A solution of acyclovir **5** (**Figure 2**) and 4-*N*,*N*(dimethylamino)-pyridine (DMAP) was added to the reaction mixture and stirred for 24 hr and then filtered. The solvent of the filtrate was partially removed *in vacuo*, and the impure solution was added dropwise to cold diethyl ether. The resulting precipitate (**6**) was filtered and dried, followed by acidolytic removal (treating the compound with trifluoroacetic acid [TFA] for 30 min at 0℃) of Bo c group to yield the desired thiazole containing amino acid esters of acyclovir (**7**).

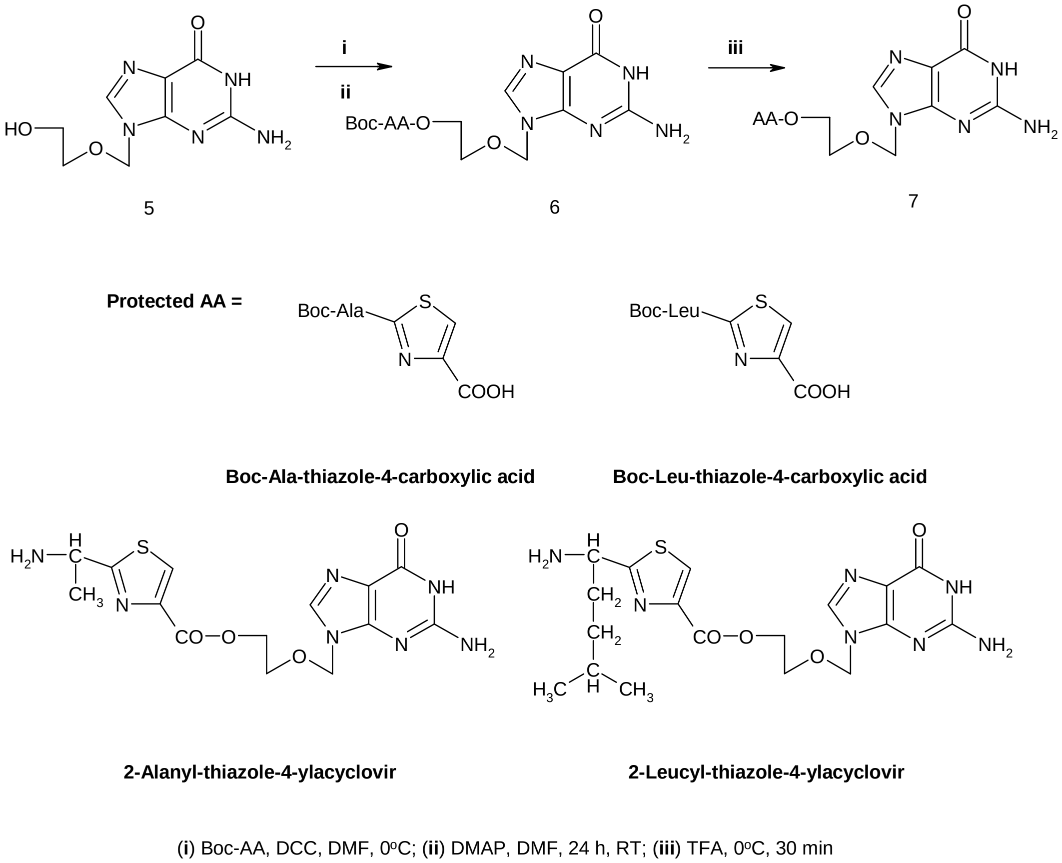


Fig. 2. Synthesis of 2-Ala-thiazole-4-yl-acyclovir and 2-Leu-thiazole-4-yl-acyclovir

2.3. Viruses and cells

The two laboratory strains - DA (HSV-1)), were kindly provided from Prof. S. Dundarov (NCIPD, BulgariaMadin-Darby bovine kidney (MDBK) cells were cultured at 37°C as monolayers in RPMI-1640 medium (Flow Laboratories, USA) supplemented with antibiotics (penicilin and streptomycin) and 10% bovine serum (NCIPD, Bulgaria). Serum concentration was reduced to 5% for growth of viruses and for testing the complexes.

Cytotoxicity assay – determination of the maximal nontoxic concentration (MNC) To compare the MNC values of substances to that of ACV confluent monolayers were covered with media modified with 40, 30 or 20μg/ml from the appropriate complex and cultured at 37°C for 96h. Samples of cells grown in test complex-free medium served as a control. The maximal concentration, which did not alter neither the morphology nor viability of the cells, was recognised as MNC.

Effect of substances on the replication of HSV. Experiments were done in multicycle growth conditions. Confluent cell monolayers were washed and infected with 320 cell culture infectious doses ($CCID_{50}$) per 0.1 ml of the appropriate virus strain. After

1h for adsorption, cells were covered with maintenance media modified with test drugs in concentrations as above. One set of infected cells served as untreated control. The effect on viral replication was determined on the 48h after culturing at 37°C by reduction of infectious virus titres as compared to that in untreated viral control. The 50% inhibitory concentration (IC_{50}) for virus-induced cytopathic effect (CPE) was determined by dose-response curve. To calculate the standard deviation of IC_{50}, each experiment was done in triplicate (for HSV-1 strain DA).

3. RESULTS

The guanosine analogues did not affect the cell morphology in investigated concentrations. Acyclovir was not cytotoxicity effect according to data in references [11]. Results from the application on their two derivatives were not unexpected - 20 µg/ml no effected cells.

The two examined derivatives and acyclovir as referent drug were applied in concentration 10, 5, 1 and 0.5µg/ml. The Ala- thiazole-4-yl-ACV and Leu-thiazole-4-yl-ACV shown insignificant effects on the herpesvirus replication – 20 and 8 % inhibition respectively (Figure 3). Whereas the referent drug inhibited the viral replication completely in same dose (10µg/ml).

These results suggest that Ala-thiazole-4-yl-ACV and Leu-thiazole-4-yl-ACV may be attractive in higher concentrations for antiviral chemotherapy obligatory with lower cytotoxicity effect in comparison with the effective nucleoside analogs.

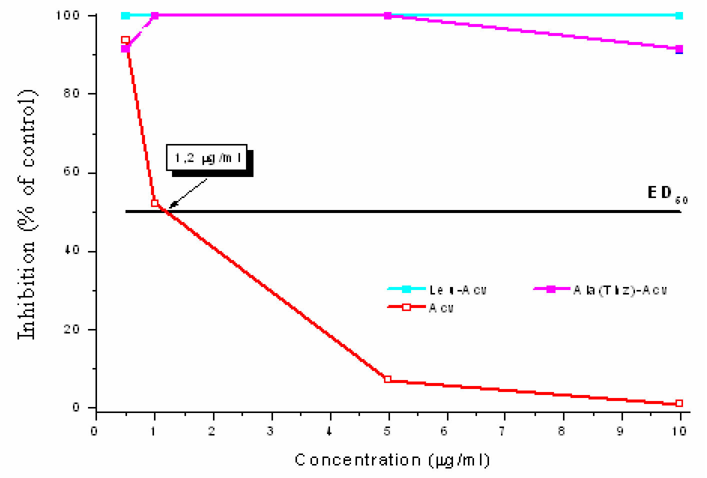


Fig. 3.

4. CONCLUSION

Design of amino acid prodrugs seems to be an attractive strategy to enhance the solubility of the otherwise poorly aqueous soluble compounds and also to afford a targeted and possibly enhanced delivery of the activedrug.

An implicit proof of this assumption lies in the fact that L-valyl ester of acyclovir (valacyclovir) shows bioavailability of 60%.

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MAN AND NATURE 6TH FORM – CHEMISTRY EXPERIMENTS

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Abstract: *A set of chemical experiments for the illustration of the educational content of the part "Matters are proposed. Matters transformation" from the subject "Man and Nature" 6th form. The aim is the teachers who do not have chemical qualification to be able to do experiments even with materials that are easy to find. A special attention is given to the virtual chemical experiment which allows demonstration even of unhealthy matters and experiments with special apparatuses. The set of chemical experiments will be useful not only for the teachers without chemical qualification but for those who will present the subject material.*

Key words: *Science, Education, Chemistry, Experiments, Teacher qualification*

По своята същност учебният предмет „Човекът и природата" представлява интеграця на физични, химични и биологични знания и осигурява възможност за изграждане на цялостен образ в съзнанието на учениците за живата и неживата природа в тяхното единство и многообразие. Този предмет е включен и в двата етапа на основната образователна степен: в начален етап (III и IV клас) и в прогимназиален етап (V и VI клас). Той е част от учебните предмети от културнообразователната област „Природни науки и екология". В настоящия етап няма учители с университетска подготовка и по трите науки – физика, химия и биология. Това налага провеждане на модулно обучение по липсващата /липсващите/ в основната квалификация специалност /специалности/. Наблюденията от това обучение за учителите, които преподават в 5 клас доказаха необходимостта от сериозна практическа подготовка за извършване на предвидените в учебната програма химични експерименти. Като се има предвид, че в болшинството училища липсва подходящо химическо оборудване, проблемът с качественото провеждане на тези експерименти става още по-сериозен. Това определи и целта на настоящата работа: да се подготви оптимален набор от химични есперименти, извършвани с лесно достъпни реактиви и подръчни средства, за раздела „Вещества. Превръщане на веществата" в „Човекът и природата" за 6 клас.

Химичните превръщания, като една от формите на движение на материята, са главен обект на изучаване на химичната наука. Затова първоначалното им разглеждане започва още в V клас в раздела „Вещества и техните свойства" на учебния предмет „Човекът и природата", като логичното му продължение е в VI клас в раздела „Вещества. Превръщане на веществата" на същия учебен предмет.

Училищният химичен експеримент е нагледно средство, чрез което едно химично или физикохимично явление се възпроизвежда целенасочено и планомерно, за да се изяснят конкретните негови признаци, особености, условия и резултати. В познавателния процес химичният експеримент играе най-често роля на източник на знания, но е непосредствено свързан и с абстрактното мислене. В тази възрастова група е от голямо значение демонстрирането на ефектни опити, предизвикващи интереса на учениците и обясняващи до голяма степен твърде абстрактния учебен материал.

Темите в раздела „Вещества. Превръщане на веществата" за VI клас са обособени в 3 части:

I. Градивни частици на веществата. Видове вещества.
1. Градивни частици на веществата
2. Химичен елемент. Прости и сложни вещества
3. Състав и строеж на веществата

II. Свойства на веществата. Химични реакции
4. Физични свойства на веществата
5. Химични промени на веществата. Химични реакции
6. Условия за протичане на химичните реакции
7. Химични свойства на веществата
8. Кислород. Свойства на кислорода
9. Химично съединяване. Оксиди
10. Получаване на кислород. Химично разлагане
11. Водород. Получаване и свойства на водорода
12. Основни видове химични реакции
13. Желязо. Свойства на желязото

III. Значение и приложение на веществата и химичните реакции
14. Веществата в природата и в практиката
15. Химични реакции в природата и в практиктическата дейност на човека
16. Опазване на околната среда

Най-подходяща за онагледяване чрез химичен експеримент е втора част, при която има възможности за организиране и на лабораторна работа за придобиване на елементарни практически умения. Особено внимание трябва да се обърне на раздела „Значение и приложение на веществата и химичните реакции", който играе обединителна функция с физичния и биологичен раздел от гледна точка на единството на неживата и живата природа.

Извършването на химични експерименти в обучението по „Човекът и природата" изисква наличие на елементарно оборудване и лесно може да се замени с подръчни средства и достъпни реактиви.

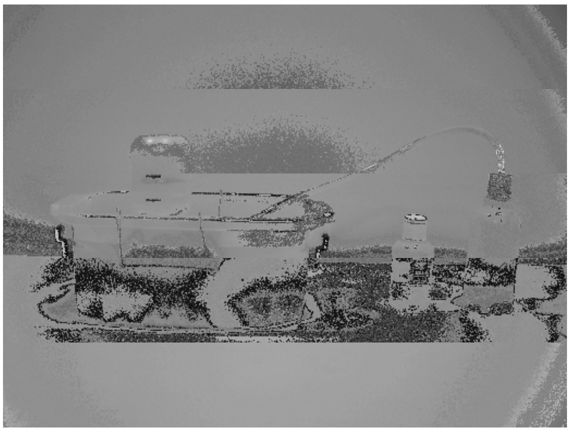
Условията и признаците за протичане на химичните реакции могат да се илюстрират с налични във всяка кухня вещества. Изискването за наличие на контакт между веществата за протичане на взаимодействие може да се демонстрира с кисело мляко и сол и сода за хляб. Насипването на сол върху лъжица кисело мляко не води до видими признаци за протичане на взаимодействие между тях, докато насипването на сода за хляб води до отделяне на мехурчета от газ (въглероден диоксид), но само на допирната повърхност между двете вещества. Разбъркването на сместа, т.е. подобряването на контакта между веществата ускорява отделянето на газ.

Признаци за протичане на химични реакции е удачно да се демонстрират с бистра варна вода, през която се продухва въздух (съдържа въглероден диоксид, който с калциевия дихидроксид в бистрата варна вода води до получаване на неразтворимо вещество – калциев карбонат). Получаване на цветна утайка може да се демонстрира с разтвор на син камък, към който се добавя натриева основа (разтвор на сода каустик). Получва се синя утайка от меден дихидроксид, която лесно може да се разтвори чрез добавяне на сярна киселина, разтвор на амоняк или по-концентриран разтвор на сода каустик. Сравнението на получените цветни разтвори с изходния разтвор на син камък и с утайката от меден дихидроксид доказва получаването на нови вещества.

Жълта утайка може да се получи при предварително поставяне на оловни гранулки в оцетна киселина, при което част от оловото преминава в разтвора под формата на йони. Когато към този безцветен разтвор се прибави разтвор на натриев или калиев йодид, се получава жълта утайка от оловен йодид.

Отделянето на газ при протичане на химична реакция също може да се илюстрира с достъпни вещества – гасена вар (калциев дихидроксид) и нишадър (амониев хлорид). Разбъркването на малки количества от двете твърди вещества води до отделяне на амоняк, който може да се докаже по специфичната миризма или чрез поднасяне към сместа на съд, съдържащ концентрирана солна киселина. Отделящият се леснолетлив хлороводород взаимодейства с отделящия се амоняк, при което се получават микроскопични частици от амониев хлорид, наблюдавани като бял дим над сместа.

Получаването на кислород и водород също може да стане с леснодостъпни реактиви и в обикновени съдове. Например кислород може да се получи освен по класическия лабораторен начин от калиев перманганат, и от кислородна вода и сок от картоф. Поради извършването на реакцията при обикновени условия и без повишаване на температурата не е необходимо използване на специален съд – достатъчна е обикновена прозрачна чаша, в която се поставят парчета картоф и се заливат с 3% или 6% кислородна вода (трябва да бъде прясно приготвена или съхранявана в хладилник). Ако трябва да се събере под вода е удачно да се вземе подходяща по размер прозрачна пластмасова кутия, в която с помощта на кламери се закрепва метална или пластмасова пластинка, на която са пробити два или три отвора. Тя служи за мост, върху който се закрепват съдовете, в които ще се събира кислорода.

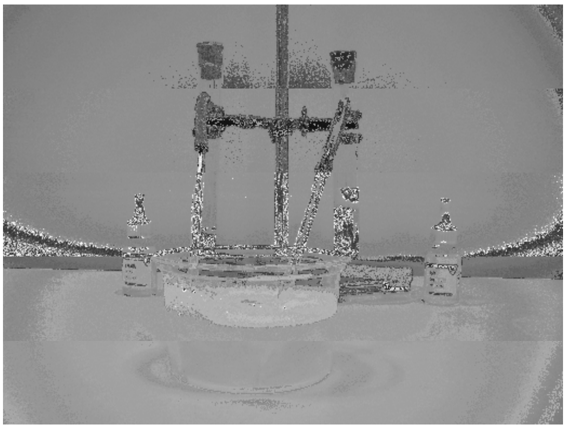


Фиг. 1. Апаратура за получаване на кислород (водород).

Получването на водород по класическия лабораторен начин изисква наличие на цинк на гранули или на прах (може да се използва и парче поцинкована ламарина) и солна или сярна киселина. При липса на киселина може да се приготви разтвор на натриева основа (сода каустик) и в него да се пуснат няколко късчета алуминиево фолио (домакинско или от капачка на кисело мляко).

За получаване на гърмящ газ, който особено впечатлява учениците, може да се използват два съда с пробити на капачките отвори, в които са прокарани сламки за безалкохолни напитки. Тези сламки се поставят в пластмасов съд, пълен с разтвор на миещ препарат. Когато в единия съд се постави кислородна вода и парчета картоф се осигурява получаването на кислород. В другия съд се получава водород по един от описаните начини. Сместа от двата газа изпълва мехурите от миещия препарат и когато към тях се поднесе запалена клечка кибрит се чува пукот – взимодействието между кислорода и водорода е съпроводено с бързо разширение и отделяне на голямо количество топлина.

Друг удачен вариант за демонстриране на гърмящ газ е събирането на водород и кислород в пластмасова банка (малка бутлка от боза или айран), градуирана предварително на 3 равни части. Банката се напълва с 2 обема водород и 1 обем кислород чрез изместване на вода. Пълната с гърмяща смес банка се внася в пламъка на спиртна лампа. Взривът е доста силен, но напълно безопасен.



Фиг. 2. Апаратура за получаване на гърмящ газ

При липса на каквито и да било условия за реална демонстрация могат да се използват възможностите на ИНТЕРНЕТ за представяне пред учениците на виртуални химични експерименти, клипове със заснети демонстрационни експерименти, анимирани модели на химични процеси или схеми на апаратури за получаване на някои вещества. Използването им в различни моменти от урока позволява да се акцентира върху най-целесъобразния елемент от продукта, да се повтаря многократно или да се илюстрира свойство, което не е удачно да се демонстрира чрез реален експеримент поради специфични изисквания за техника на безопасност, отделяне на вредни вещества или липса на изходни реактиви.

ЗАКЛЮЧЕНИЕ

Анализирано е учебното съдържание по учебния предмет „Човекът и природата” в 6 клас и е съпоставено с изискванията на ДОИ за прогимназиална степен и учебната програма. На базата на този анализ е направен подбор на химични експерименти, подходящи за онагледяване на учебния материал. Предложени са варианти на демонстрационни експерименти, изпълними с лесно достъпни материали. Посочени са някои специфични технически и методични особености за извършването и използването на експериментите в контекста на конкретното учебно съдържание.

Направен е подбор на виртуални химични експерименти (събрани в отделен CD), които могат да се използват самостоятелно или в комбинация с лабораторни и демонстрационни експерименти. Те могат да се намерят в сайта на катедра „Химия" при ПМФ на ЮЗУ.

Посоченият набор позволява учителят да подбере най-подходящите за учениците и условията експерименти, свързани с онагледяване на учебния материал в раздела „Вещества. Превръщане на веществата" на учебния предмет „Човекът и природата" в 6 клас. Не е задължително да се използват много на брой опити, а целенасочено и методически издържано да се селектират за най-ефективно достигане до крайната цел – постигане на заложените в Учебната програма очаквани резултати от учениците.

Интересът и към химичния експеримент в тази възраст може да се изиползва като една добра предпоставка за мотивиране и насочване към природните науки при бъдещата професионална ориентация и реализация на учениците.

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ПОСТНЕКЛАСИЧЕСКИ ПРЕДСТАВИ В ОБУЧЕНИЕТО ПО ОБЩЕСТВЕНИТЕ НАУКИ (СИНЕРГЕТИЧНИ АСПЕКТИ)

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Динамиката на процесите в обществото през последните 20 години в страните в „преход" от Източна Европа се характеризира главно с процеси на дестабилизация и разрушаване на старите обществено-политически и икономически процеси. По същество тези процеси станаха част от специфичната самоорганизация на нови отворени сложни социално-политически структури, както и на съответни изменения в образованието и науката на XXI век. Новите методи на изучаване, изясняване на тенденциите, на основни параметри на регулиране на тези процеси и структури, на стабилността и предвидимостта на тяхното развитие изискват и качествено нови подходи в обучението.

Една от важните задачи в тази насока е осъзнаването на идеите и понятията на новите алтернативни методологии, на промените в методиката на обучение по обществените дисциплини – най-вече поради бързото навлизане в тях на тези идеи и необходимостта от създаване на условия за правилното усвояване на методите на така наречените „постнекласически” науки[4]. Такива са синергетиката и нелинейната динамика [2], чиито методи адекватно описват така наречените „фазови преходи” от типа хаос-ред, особености, бифуркации, атрактори, „управляващи параметри”, устойчивост (стабилност на динамиката) на развитието – в това число на обществено-политически процеси в най-новата история на човечеството.

Концепцията за устойчивото развитие, приета от ЮНЕСКО през 1992г., докладите на Римския клуб от почти 40 години насам (или в по-точен за стабилна динамика на развитието) изисква адекватни на новите условия методи на обучение, достъпни методи на представяне начини, методи и понятиен апарат - с оглед и на необходимостта на устойчива динамика на развитие на образованието (за такова стабилно или устойчиво в този смисъл развитие). Когато става дума за „устойчивост”, следва веднага да се уточнява правилно целият комплекс от относителни понятия и взаимовръзки в постнекласическата методологическа парадигма. Така и представите за *устойчивост* са също относителни и динамични.

Всичко това налага и съответните промени на методиките на обучение, които все по-често ще разглеждат процесите нелинейно и нееднозначно. Тоест следва да създават основата за изработване на „нелинейна култура” на мислене. Най-малкото защото е известно, че *пълната устойчивост противоречи на развитието,* тоест такива системи не съдържат в себе си източници на развитие. Обновяването особено на методиките на обучение по историческите и социално-политическите науки предполага разработването на модели, задачи, планове, методически единици и пр., които да съответстват на новия тип нелинейна култура, на новите нелинейни науки за сложните отворени самоорганизиращи се системи (по същество това се отнася и за самите съвременни социохуманитарни науки). Динамиката на такава система (например на една обществена система, с която се занимават и историци, и политолози, и икономисти, и културолози и т.н.) предполага специфично съотношение между устойчивост и неустойчивост. При това такова, което системата да „живее” и се развива – именно чрез обмяна на информация, на енергия, на човешки потенциал и т.н.

Важни примери в тази връзка, които често се налагаше да привеждаме дори преди, а по-късно - и в началото на силно хаотизираните преходи у нас и в Източна Европа (в края на 80-те години на XX век) бяха свързани с имената и новите идеи на представите за самоорганизация на системите и тяхната устойчивост [1]-[7], [8], [12], [13]. В тази връзка започнахме работа и по наш международен проект от февруари 1986 г. по интегративно-синергетични аспекти на глобалните проблеми на

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[4] В частност, ако сравним представите например на историците за нелинейната наука за сложното (като нова синергетична парадигма), от една страна, и от друга проанализираме техните спорове със спекулации от типа на така наречената „Нова Хронология” (за датировките на историята), то се оказва, че историците по-добре са информирани за схемите на една спекулативна нова „хронология”, отколкото за идеите и понятията на новите нелинейни научни методи, благодарение на които за кратък срок са получени изключителни резултати в много науки. При това за много от историците е трудно аргументирано да опровергават неприеманата от тях „нова хронология”. Но пък продължават да отказват да се запознаят с идеите и методите на науки като синергетиката, съответно да правят съпричастни и своите студенти. В началото на XXI век образованието по природните и на социохуманитарните науки следва да има по-ясен отговор защо се получават такива различия и какво да се промени в него.

човечеството и изменение на методиката на обучение по социохуманитарните науки. Ръководството на съвместния проект се осъществяваше от ръководителя на *катедрата по философия* на МГУ на чл.кор. С. Мелюхин, при участието на акад. Ив. Фролов, на акад. Н. Моисеев [7], на проф.Зотов, на проф. Бестужев-Лада и много други учени от различни области. Тук ще подчертаем една мисъл на акад. Н. Моисеев за абсолютизирането на методологията и на „затварянето” тогава на много от обществените науки в страните в източна Европа. Тя напълно се отнася и за инерционността в промените и устойчивостта на методиките на обучение, свързани с онази устойчивост и на самите обществени науки (или по-точно с идеите на техни представители), които поради такава „устойчивост” днес се оказаха неспособни дори да удържат своя предмет в средата с поливариантни икономики, идеологии, политики и пр. , а именно, че „..*устойчивостта, доведена до своята граница, прекратява всяко развитие.Тя противоречи на принципа на изменчивостта. Прекалено стабилните форми – това са „тупикови” форми, еволюцията на които се прекратява. Прекомерната адаптация... е също толкова опасна за съществуването на вида, както и неспособността към адаптация*”[7]. Докато идеите на синергетиката, нейните методи водят към осъзнаването на сложността и нелинейността на света, в който живеем, свят в който протичат преходи и самоорганизация от типа „хаос-ред” и обратно.

В този смисъл и въвеждането и осъзнаването на понятийния апарат на синергетиката в много от темите на различните обществени науки дава адекватна съвременна интерпретация на детерминизма и индетерминизма. Така както квантовомеханичните идеи и методи, както вероятностно-стохастичните представи за света, както релативистичните представи и т.н. в началото на XX век създадоха новите научна представа за света. Те бяха в основите на създаването и на новите технологии на бъдещето, на много нови научни дисциплини (в това число чрез осъзнаването и използването на новите методологии и методики на обучение). Сложните процеси на нелинейната динамика, изучавана от различните науки на XX век, както и динамиката на самото образование в края на XX век, на самите информационни технологии родиха и науки като синергетиката. При това се стъпи и на идейната основа на много от научните идеи на XX век – на синергетизма в лазера, на теорията на неравновесните дисипативни структури и т.н. Те и подобни на тях дадоха и нова нова интерпретация на детерминизма и индетерминизма. В това число могат да се посочат и такива примери като модела на така наречения от Е.Лоренц през 1963 г. «страннен атрактор» за нестабилното неравновесно поведение, в което малките въздействия (или изменения на началните условия) могат да доведат до огромни изменения. В този смисъл съвременните обществени науки вече трябва да бъдат преподавани така, че тези идеи да са в реално съответствие с техните изводи и понятийния апарат.

Сред създателите и първите популяризатори на новата „Х-наука” за самоорганизацията на сложните нелинейни системи са такива като кръстникът на синергетика Хакен [13], Пригожин, на плеяда руски учени като Манделщам, Адронов, Колмогоров и много други по нелинейна динамика на трептенията, на Арнолд, на френския учен Р. Том, на Р. Постон и т.н. Що се отнася до признанието на ролята и мястото на нелинейната теория на самоорганизацията в обществените науки ще отбележим например, че още 1987 г. едновременно с другите си високи научни длъжности и ангажименти, Иля Пригожин става директор и на института за *социални науки* L’Ecole de Hautes Etudes, Франция. Това преди всичко е реална оценка за ролята на новите евристични и относително доста универсални идеи, за ролята на новите и мощни методи за изследването, разбирането и възможностите за предвиждане на социалните процеси. Именно тази ефективност и многопосочност на приложението на методите на нелинейната наука дават пълно основание на учени от множество различни направления да я възприемат като нова и важна *синергетична* парадигма, като основа на нов тип интегративност и интердисциплинност, родена и прокарваща си път през не малко научни спорове от 70-те години на XX век насам. В началото синергетиката или теорията за самоорганизацията на сложните отворени неравновесни системи с различна природа повече стихийно си пробива път и място в науката и образованието. Това става предимно в природонаучните области, а през 90-те години започва бързото (и *не винаги коректно*) навлизане на тези идеи и в широкия спектър от обществени науки. Техните представители донякъде са били, а и още са затруднени в обяснението и моделирането на новите реалии в глобалния свят на свръхинформационните технологии. Сферата на образованието все по-силно се нуждае от нови и алтернативни подходи за *осъзнаване на общото в динамиката* на съвременната научна картина.

Човечеството премина в третото хилядолетие с качествено нови идеи, родени от обществено-политическото си (и образователно) развитие през XX век. Но и Хакен, както и Пригожин и др. са доста внимателни в използването на синергетиката. И преди всичко - *доколко* картината на самоорганизиращо се общество в този „век на бифуркации" може да бъде „*изведена*" от теорията на самоорганизацията. Хакен сочи: „ *...аз съм длъжен да оставя открит въпросът дали е реализуема такава форма на общественото устройство с нейните детайли или тя е безусловно желателна. Във всеки случай исках да дам стимули за по-нататъшното движение на мисълта и за продължаване на дискусиите по тези фундаментални аспекти, когато отново и отново става дума за това как да се балансират интересите на отделните хора и интересите на обществото*".

Така или иначе обществото, политиката, а и самите обществени науки (като сложни неравновесни нелинейни системи) имат не по-малко основание да бъдат осмисляни професионално чрез синергетичните идеи. Защото синергетиката се разглежда именно като наука за самоорганизацията на системи, които са преди всичко големи, сложни, отворени, съществено нелинейни, силно неравновесни системи. И в тях протича самосъгласувано-кооперативен процес на самопроизволно качествено изменение на системата, насочен към варианти на самоподреждане (увеличаване на вътрешната организация). Внимателното вглеждане в общото на социално-политическите науки и образованието по тях днес показва *нови и не винаги ясни представи* за своеобразните взаимодействия и иерархии (на интегриране, синергизъм), както впрочем и към разграничаване в различни науки:

1. Днес като цяло обществените науки все още рядко „виждат" естественият „*синергизъм на науките" и* дори на тези, в недрата на които се роди синергетиката (в широк смисъл). Не се осъзнава смисълът на използвания понятиен апарат, още повече, че той не е изкуствено създаван, **а** *идва от* важни предшественици и *извори на синергетиката*. Такива са и кибернетиката; физиката – лазерната физика на Хакен, неравновесната термодинамиката, физикохимията, теория на дисипативните структури, физика на отворените системи, физиката на фазовите преходи); от теорията на динамичните системи, нелинейната механика на трептенията, теорията на катастрофите, теорията на фракталите, астрофизиката; и биофизиката, биохимията и теорията на информацията, информатиката и т.н. Общите подходи и специфика на тези съвременни направления са свързани и с „интер-„ или „мулти"дисциплинарните възможности на различни математически теории и модели - качествената теория на диференциалните уравнения, на особеностите и бифуркациите, вероятностно-стохастичните теории и представи, и др., информатиката и съвременните възможности на информационно-компютърните технологии, новите възможности на РС-визуализацията на фазови портрети на сложните решения на нелинейните диференциални уравнения, описващи еволюцията на съответната система и сложни нелинейни модели (като моделите на Е. Лоренц от метеорологията и топлинната конвекция от типа на странния атрактор и много други) и т.н.

Във връзка с горното следва да се отчитат и взаимодействащи направления, развиващи и допълващи чрез синергетиката. Това са съвременната политическа глобалистика, глобалния еволюционизъм (Е. Ласло, Н. Моисеев и др.), дали тласък и на политическата регионалистика, екополитологията, коеволюцията на човека, обществото и природата, на синергетичните оценки за единството природонаучното и хуманитарното знание, съзвучието с глобалните проблеми на човечеството и др. Това е и теорията на автопоезата (У. Матурана, Ф. Варела [6] и др., представящи оригинална концепция, ознаменувала появата на нова междудисциплинна научна парадигма и породила дискусията за радикалния конструктивизъм [2]). Това е и теорията на сложността (М. Гел Ман и др) и др.[5]

2. През последните години объркващо навлизат нови понятия и „стесняващи" определения за синергетиката - предимно от страна на представители на обществени дисциплини. Стана популярно да се създават и обсъждат *множество „синергетики*". Така, от една страна допреди 15-20 години (а не рядко и досега) „синергетиката" бе посрещана едва ли не със същите начални оценки, както при появата преди това на други такива качествено нови науки (например кибернетиката, представяна като „лъженаука", същото бе и при генетиката и т.н.) - при това най-вече в „полето" на обществените науки. От друга страна, през последните години пък именно сред тези науки се появиха не съвсем коректни представи за *множество синергетики*. Така не става ясно каква е принципната разлика между теоретичните основи на „различните синергетики", такива като: социална синергетика, лингвистична синергетика, *икономическа синергетика* (защо не и *политическа синергетика*) и др. Тоест едва ли не се появяват вече толкова „*различни синергетики*", колкото са и различни синергетици в света. Явно се прави грешка при разграничаване на единния подход (ако тези „синергетики" не са еквивалентни на смисъла на обърнатите „двойки" (не „икономическа синергетика", а „*синергетична икономика*"- Занг, Пу и др.). Тоест, забравя се, че се „търси" ролята на синергизма *в предмета* (процесите, взаимодействията и пр., изучавани по единен начин във всяка от тези дисциплини, а не обратно, което обърква логиката на търсенето). Малко по различен е случаят, когато говорим за „образователна синергетика", когато се имат предвид принципно новите възможности, методи и пр., виждани през призмата на новия съюз „синергетика и образование". Става дума за спецификата на синергетичното взаимодействие в образованието (а не„разделящо" и „определящо" отношение на една наука към друга). За синергизма на психологически, педагогически, управленски, физиологични, информационно-компютърни и комуникативни технологии, на

[5] Ограничеността на изложението тук не позволява да поясним липсата на изисквания към системите от параметри, критерии, типове и принципи на самоорганизация, които още не удовлетворяват свойствата пълнота, непротиворечивост и независимост (доколкото е възможно да станат такива и да се стигне до необходимата математическа формализация, което е дългосрочна задача и за синергетични колективи от учени – подготвени за работа с представители на природните науки)

икономически и други фактори и „управляващи параметри”. Същите дори при по-малки, но правилно насочени „кохерентни” усилия водят до качествени изменения (фазови преходи като тези при ученето, запомнянето, логическото осмисляне, разбиране, ефективното приложение на съответни знания). Понякога се налага да се разгледат различни аспекти – като синергетика на образованието, синергетика и образование, синергетика в образованието и др., Те навлизат в специфични страни на синергизма, но *не като различни синергетики*.

При това често се губи и представа какво е специфичното синергетично „нелинейно сумиране” (усилване) при даден ефект. В частност трудно се обяснява защо резултатът от дадено физично (да не говорим *за политическо*, икономическо, образователно) *синергетично взаимодействие* е от типа: „1 + 1 дава 5” (грубо и образно казано). А това е част от спецификата, заради която понятието синергизъм е използвано още преди 160 г. от физиолога Ч. Шерингтън - за съгласуваното действие и управление на мускулните влакна от гръбначния мозък. Аналогично то се използва и във фармацевтиката - за „умножаването” на ефекта на едновременно използване на някои лекарства - тоест за получаване на ***свръх***сумарни ефекти и пр. Така представите за „различните” синергетики въобще не поясняват каква е същността и спецификата на това „различие”. Още по-малко те обвързват този „нелинеен ***свръх***резултат” с това какъв точно понятиен апарат се занимава с еволюцията на подобни системи и техни геометрични и PC- представяния, или с това какво е характерното за един фазов преход, какво (най-общо) представлява всъщност едно „нелинейно диференциално уравнение” и пр.

Представителите на обществените науки разбира се не са длъжни да са специалисти и в тези природоматематически области. Но за да погледнат синергетично към своя предмет на изследване (а и към образованието и особено по обществено-политическите науки) се предполага допълнително образование (самообразование), работа в интегративно-синергетични екипи и пр. Но днес дори не се вижда и изчезващото присъствие в образованието по обществените науки дори на уводни съвременни курсове по природни науки или поне на курсове по съвременни концепции на природните науки (във висшите училища). Така не се обръща внимание и на реалното редуциране (или отсъствието де факто) на осъвременени в тази насока курсове по философия, политология, психология, педагогика и др., насочени към специалностите по природните науки.

3. След изминалите 90 години след гигантските „флуктуации” и „бифуркации” през 1917 г. в политическия живот в Русия и страните в „преход” в Източна Европа. Още в началото на сложния преходен период след 80-90-те години на XX век започна бързата институционализация в тези страни на социално-политически науки (с формирането на новите катедри, с новите специалности и учебни програми). Качествено нова насока се наблюдава в тенденциите на развитието на обучението по всички социохуманитарните науки. При това днес статистиката в големи страни като Русия, в Украйна например отчита много бързо нарастване на студентите и дисертантите по политически науки. Например от 1995 до 1997 г. е регистрирано *над тройно такова нарастване* (надхвърлящо динамиката на нарастване по други „пазарно” търсени специалности по обществени науки). Взривообразно расте през 90-те години на XX и началото на XXI век и броят на публикациите за приложенията на синергетиката в областта на социохуманитарните науки и образованието. Този тип динамика е известна в нелинейно-синергетичените модели като нелинейния синергетичен „режим с изостряне” (blow-up solutions) с хиперболично нарастване[5]. Друг е въпросът доколко *тези идеи отразяват адекватно специфичното* в такава динамика на обществено-политическия живот на нашето съвреме.

В този смисъл виждаме и една от насоките на образованието по политическите и другите обществените науки - именно все по-задълбочено да обсъжда новото място, роля, тенденции, взаимодействие на съответната научна дисциплина, както и нейната самостоятелност, особеност и вътрешна *цялостност.* В тази връзка е важно и в образованието да се разделят по-ясно науките като преподавани курсове (с помощта на обновявани адекватно методики) и като научните изследвания при бързи промени и в двете насоки. Освен това често не се различават спецификата и промените например в политическата наука като преподаваем курс от собствено политологическите изследвания. Често последните в много от страните в преход (поради хаотизиран политически процес на прехода и повишено им търсене особено в медиите) претендират за научност на изследванията, оставайки си при това в доста от случаите частни разсъждения за политиката, без да е усвоен и използван инструментариума на съвременния политолог (в т.ч.: теорията на системния анализ, теорията на приемане на решения, теорията за обществения избор). При това за разлика от западните школи и центрове в Източна Европа през този преход бе трудно да се види системно обучение и използване на математически методи на анализ на данните, методи на моделиране, на социалната статистика, на корелационния и клъстерен анализ, на теория на игрите и др.

За съжаление и след изключително удачната и навременна книга М. Бушев [2], издадена 1992 г. (скоро след началото на социално-политическите промени у нас) , академичната общност като цяло остана индиферентна към необходимостта от нов синергетичен съюз не само на обществените и природните науки, но дори и към синергично взаимодействие между природоматематически науки, породили новата нелинейна наука. И сега необяснимо отпадат дори от физикоматематически специалности базови дисциплини като теория на вероятностите, превръщат се в „избираеми” (тоест неизбираеми) и със силно занижен хорариум математичните методи на физиката, статистическата физика и термодинамика и др. От друга страна, вниманието на студентите природонаучни специалности не достатъчно е насочено към исторически, социално-политически, философски, и др. дисциплини. Така вече над 15 години откакто синергетиката показва активно присъствие в изследванията и обучението по обществените науки, образованието у нас не се заема със задълбочено обсъждане на синергизма на науките.

Въпросите от този характер днес *имат не само риторичен характер*, особено що се отнася до спецификата на проекциите на тези идеи им върху съвременните исторически и социално-политически науки в страните от Източните Европа именно през последните години. Тоест това става именно в същия период откакто се формира системата на Европейския съюз. И едновременно през този период във всички области активно търси своето място и съвременната *синергетична парадигма*.

Така синергетичната парадигма намира своята реализация и в съвременната педагогика. Засега сме свидетели на хаотично раждане на своеобразна „синергетична” педагогика. А това ще се реализира чрез съответни промени в съвкупността на теоретични, методологични и други установки, апробирани от научната и педагогическа общност (самоорганизираща се също по нелинейните синергетични механизми) при променящите се етапи на развитието на педагогиката и т.н. Именно такъв тип синергетична педагогика може да съответства на педагогиката на постиндустриалното (информационното) общество. Тя се формира в съвършенно нови условия и развитие на новото общество,при бързо изменящото се образование – особено при динамиката на промените на обществените науки.

Изводите от представеното по-горе са свързани преди всичко с налагащите се промени и в методиките на обучението по обществените науки – тоест при новите условия на промените в информационното общество и образованието. И тук не става дума само за бягането от типичните идеологеми от преди края на 80-те години на XX век. А за методика на обучение по обществени науки, съответстващо на ускорените промени и на нелинейната синергетична парадигма и в образованието [9]-[11].

А и всичко това не е прост акт на написване на поредни пособия и методики, а следва да се прави от синергетични колективи в академичната общност с оглед промените, идеите и мощта на интердисциплинарните и синергетични подходи. Особено важно е това да става успоредно с промените в учебната документация (планове, програми), свързани със синхронизацията с изискванията на Болонския процес (които сами по себе си са синергетични управляващи параметри в управлението и регулирането на новото отворено образование на бъдещето). Явно че академичната общност тръгва в тази насока засега недоподготвена за необходимото за целта изследване и описание на конкретните начини, техники на педагогическата дейност в отделните образователни процеси отчитащи спецификата на синергизма на науките и реалната самоорганизация на образованието в условията рязкото нарастване на обема и сложността на информацията особено във висшето образование. В частност липсва ясна представа и системна подготовка у голяма част от преподавателите по обществените науки - за новият понятиен апарат, касаещ и нелинейните методи на синергетиката, за съответните промени и в методиката на педагогическото изследване. Така преди всичко съвкупността на начините на организация и регулиране на педагогическите изследвания, на реда на тяхното прилагане и на интерпретацията на получените резултати за достигането на определената научно-методическа цел трябва да се изменят от гледна точка на новата „нелинейна култура на мислене”. Чрез нея следва да са изясняват бифуркациите, описанието на механизмите на качествените промени (невъзможни при прилагането на класическите линейни методи), бързите промени в обучението при усилената обратна връзка и индивидуалното регулиране на PC –интерактивното взаимодействие и участието на „машината” в самоорганизацията на обучението в зависимост от индивидуалните възможности на обучаемия и пр., обезпечаващи адекватни индивидуална скорост и управляващи параметри в „синергетиката на ученето”.

В подобни насоки следва да се очаква общността да изяснява и прилага синергизмът в цялостната организация и осъществяване на учебно-познавателната дейност. Преди всичко това се отнася и за промените организацията на взаимодействие с източниците на информация и възприятие (лекции, диспути, обяснения, и пр., които често ще следва да отчитат дистанционните ИКТ и ТВ технологии; за промените в нагледните методи (илюстрации, и пр.) и промените в практическите методи (семинарни и практически упражнения, задания с PC-визуализации, мултимедийни методи и демонстрации при възможността за тяхната честа и бърза повторяемост колкото и където е необходимо на обучаемия). Аналогично това се отнася към промените в методите и логиката на мислене (индуктивните и дедуктивните методи на обучение), вероятностно стохастичните методи и промените свързани с методите на регулируемия динамичен хаос. В тази насока могат да се разглеждат и синергетични аспекти на активността на познавателната дейност – на активното възприемане, запомняне, възпроизвеждане на учебна информация.

Последните две десетилетия показват в различни области, че интегративността, междудисциплинността и синергетизмът в динамично променящите се системи на науката и образованието са нов нов мощен потенциал, налагащ бързо осъвременяване на методиките на обучение по историческите и социално-политическите науки във висшите училища през XXI век.

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FOSTERING YOUNG FEMALE SCIENTISTS IN THEIR ACADEMIC CAREERS: THE ADVANCE PROJECT

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***Abstract**: The paper deals with the gender equality issue in the field of science at European level. The objective of this text is to promote the women in science issue and contribute to finding „mechanisms for involving women scientists more actively in research management and policy making". It presents information about good practices and provides ideas for academic initiatives which aim at supporting female scientists in their career development. Such initiatives in our case is a Summer School and a Mentoring and Coaching Program.*

***Keywords:** European research, science policy, female researchers, human resources for science, scientific career development, advanced training, summer school, mentoring and caching, transfer models.*

INTRODUCTION

The process of European integration focuses the attention of different communities over the continent upon challenges and issues of key significance and common value. It mobilizes strengths and organizes the efforts through creating platforms and frameworks as well as by establishing programs and projects which aim at combining resources to solve crucial problems together. Within these, individual members (people, institutions and organizations) are able to share ideas and make use out of the abundance of cooperation activities.

One of these initiatives of the European Union is the creation of the European Research Area which was decided in 2000 and has been developing intensively ever since. It's been regarded as a means of establishing a unified field over Europe which should enable researchers to move, interact, and share knowledge and experience. It aims at supporting best research throughout Europe, coordinate scientific programs and promote development of strong links with partners around the world thus contributing to global development and benefiting from the worldwide progress of knowledge. The Research Area has been also intended to inspire the best talents to enter research careers in Europe.

Recognizing the place and role of women scientists, the European Commission has adopted an Action Plan on Science and Society one of which actions is "Promoting gender equality in science in the wider Europe". The latter coincided with the Commission's initiative to "mobilize women scientists to enrich European research" (EC, 1999). For the same reason, the Commission also set up a group of experts (consisting of senior scientists from various disciplines, representing academies of sciences, universities, research institutions, administration and business) in 2002 responsible for implementing the action. This body is known today as the ENWISE Expert Group that stands for – Enlarge Women in Science to East.

Since then the experts' major task has been to study and report on the situation of women scientists in ten ex-socialist countries (including Bulgaria). In its report from 2004, the Expert Group states that in the ENWISE Countries: "Women are still under-represented at the top positions in academies of sciences and in universities. Women constitute the majority of teaching staff (54%), but tend to be concentrated in the lower academic positions. Furthermore, despite the fact that women's participation among university staff is similar to their presence as researchers, men are three times more likely to reach senior academic positions than women" (EC, 2004: 7). In addition, it specifies that: "The prospects of young female scientists are very bleak due to the unavailability of funding, the rigid patterns of promotion and recognition, and the lack of appropriate welfare policies" (ibid.).

All these findings call for urgent attention towards the young female researchers and elaboration of policies and measures to foster their scientific careers. At European level the Sixth Framework Program (2002-2006), based on the key concept of building the European Research Area, established a section "Women and Science" which aims to "boost gender equality in research, through stimulating the participation of women in science and technological development; and fostering the integration of the gender dimension throughout European research" (Work Program: 2005-2006, Science and Society). By its Women and Science Call for projects it encourages and supports activities which focus on "supporting or linking initiatives designed to promote women in decision-making and policy shaping positions, such as: networking; mentoring in career development; role models; specific training and coaching programs; fast-track systems" (ibid.).

WOMEN IN SCIENCE IN EUROPE AND BULGARIA

The situation in Europe

The position of women in society has changed considerably over the last decades. This trend raised much greater consciousness of the need for women to be encouraged to aim for careers which were not previously opened to them and especially for managerial positions in their proper field of professional development. The situation with women in the domain of science is much the same as can be seen in Tab. 1.

As a matter of fact there are lots of studies on the women in science issue which show that the situation in Europe is rather complex. The share of women in the total of researchers in 2001 was below 50% with the exception of Latvia. In the older EU (15) member countries it's less than one third, and the female share of the population of researchers in Austria and Germany was extremely low. At the same time, Bulgaria was among the countries with the highest share of women in scientific careers even though their representation on managerial positions was still not so high.

Tab. 1: Female researchers as % of all researchers in EU

Latvia	52,7	Ireland	29,4
Lithuania	47,0	Cyprus	29,3
Portugal	46,6	Finland	29,1
Bulgaria	**45,5**	Norway	28,3
Estonia	43,3	Denmark	28,0
Romania	42,8	Italy	27,9
Greece	40,9	**EU – 15**	**27,2**
Poland	38,1	France	27,5
Slovenia	36,8	Czech Rep.	26,8
Acc. countr.(3)	35,5	Slovakia	24,0
Spain	35,4	Switzerland	21,2
Iceland	34,6	**Austria**	**18,8**
Hungary	33,0	Germany	15,5

Source: DG Research Key Figures 2003-2004

In its report from 2002 the EC provide the following worrying data: "In the EU Member States in 2000, women made up 41% of undergraduates in science, mathematics and computing and 20% of those studying engineering, manufacturing and construction subjects. In science, their share is highest in life sciences (50%) followed by mathematics (30%), physical sciences (27%), engineering (20%) and computing (19%) (EC, 2002).

Women remain seriously under-represented in the science and engineering fields of study, especially in engineering where they represent only 22% of all graduates in Europe in 2004. As researchers, women are particularly under-represented in the business sector, as well (EC, 2004). The situation is the same in industrial research where the position of women "faces the same problems, as old fashioned ideas and practices still impede women's careers in industrial research. At present women constitute only 15% of industrial researchers in the EU (EC, 2002).

Grounding on the data accompanied by proper analyses, the EC came to a very important conclusion: "women are widely recognized as being an important resource for European research", they "have a huge potential for the future of research in Europe", but also that "their huge potential is under-exploited". (EC, 2004)

The under-representation of women in science and especially on managerial positions in the sector results from many different factors (historical, social, cultural, professional or is caused by specific career paths and models) which vary from country to country even though there are some common ones. In any case, the investigation on the nature of these factors is crucial for the identification of key problems and elaboration of proper strategies and initiatives aimed at encouraging the participation of women in science by the university governing bodies.

GENDER ISSUE IN BULGARIAN SCIENCE

The share of the female researchers in the public funded R&D sector in Bulgaria was 48,8% in 2001 and 49,3% in 2002 which was among the highest in Europe.

In her recent publication Bulgarian researcher Nikolina Sretenova (Sretenova, 2006) is exploring the extent to which the gender balance issue is present in the official documents and regulations concerning science in Bulgaria. These are: the Innovation Strategy of Republic of Bulgaria and the Measures of Its Implementation (2003), the National Strategy for Scientific Research (Project) for the period 2005-2013 (2004), and the Law for Scientific Research Promotion (2004). They were elaborated in response to the EC's criticism of some legislative "white spots" discovered in the Bulgarian R&D sector during the accession negotiation process (2000-2004). The analysis of the documents made by Dr. Sretenova makes her to conclude that "these three official documents are not gender sensitive". She furthermore concludes also that: "Unfortunately, the EC policy for mainstreaming equal opportunity or achieving gender balance in scientific research financed from public sources was not considered, nor were the ideas incorporated into the provision of these documents". (ibid.: 228)

Actually, there are just two paragraphs only in the Strategy concerning the women in science issue. They say literally the following: "Regarding the human resources in the European Union as a whole, there is a big unexploited potential, connected with the women participation in the research and development sector. While in the EU an average the total number of female researchers is about 27%, in Bulgaria it is 45%. It means that there is a relatively balanced participation of women in research in the country in spite of various defects like their small share in managerial positions" (MES, 2004: 11), and: "There are no serious problems with the share of women in the total of students which study this group of disciplines – natural sciences, mathematics and engineering." (MES, 2004: 12)

THE AUSTRIAN EXPERIENCE

As we have already seen in Tab. 1, Austria is among the countries where the share of women researchers in the population of scientists is considerably low. Some specialized surveys have shown that the gender segregation at Austrian universities still exists and it is especially strong at the senior management level.

In order to promote women in the field of science and research diverse activities have been undertaken in the country. Evi Genetti, in her recent publication, describes the pioneer movement towards fostering women academics in their professional development at the University of Vienna. A career-planning course, titled "Recognizing potential. Developing visions. Designing the future. Women in academies plan their career", was first organized. After that scientific writing workshops for female students as well as gender specific training for teaching staff were offered (free of charge) and carried out by external coaches. These were implemented in a form of a project which major aim was to support female students in overcoming difficulties, working on their theses for the first and second degree.

As Genetti reports, one of the most successful projects implemented in Austria was the "Mentoring University Vienna" (Genetti, 2006: 193). It was a mentoring program for female academics and scientists at doctoral, post-doctoral and habilitation levels. The program was organized in a form of a interdisciplinary training with some cross-gender activities.

Formal mentoring program for women academics in Europe were first set up around the end of the 1990s. They proved to be of great value for women when supported in crucial stages in their scientific careers (Leemann, 2000).

Here we have to clarify the two basic terms: "mentoring" and "coaching". According to the definitions, accepted in the ADVANCE project (that we shall discuss further), "mentoring is a long term relationship that has both, a personal and a professional dimension" (ADVANCE, 2006). Usually, it is established between two persons (a mentor and a mentee) and aims at promoting the mentee in terms of career development, networking, organizational know-how, etc. within the academic and industrial research context. In comparison, "coaching is perceived as a short term relationship" which "provides a special focus on certain professional or personal issues". It can be organized both individually (individual coaching) or in small groups (group coaching) and aims at "a quick and focused collaboration between the coach and the coachee, the former supporting the latter in developing and using her own skills" (ibid.).

The mentoring program of the University of Vienna supported around 40 mentees planning their scientific careers and chasing professional goals. A complementary aim of the project was to provide junior women access to formal and informal scientific contacts and networks.

The mentoring project become gradually a best practical experience and served as a model for other academic mentoring program in Austria.

THE 'ADVANCE' PROJECT

ADVANCE is an acronym which serves as a title of a project under the 6^{th} Framework Programme and stands for "Advanced Training for Women in Scientific Research" (ADVANCE, 2006). It addresses the issue of the gender equality in science and research and intends to contribute to improving the situation by supporting female scientists from several EU countries. More concretely, the project intends to promote the participation of women in science and research decision making and policy definition. So, as a result of the project, the participating female scientists are expected to obtain a broad range of research and management skills and tools relevant to their career advancement.

The project was initially designed by a team of experts from the Danube University Krems in Austria and then discussed and agreed on with 5 other partner universities from 4 European countries (Bulgaria, Poland, Finland and the Netherlands) which got involved in the final proposal. It was successful and got the required funding by the Commission. The South-West University "Neofit Rilski" has been the Bulgarian partner in this project.

OBJECTIVES OF THE ADVANCE

As it has become evident, it is crucial nowadays to offer women scientists easy access to training, expertise, networking and provide role-models in order to alter the existing situations where they are under-represented and their potential underexploited. That's why the ADVANCE programme proposes different forms of support in which the participants are able to learn and enhance their abilities, build supporting networks and receive the encouragement they need to further their professional development.

One of the major goals of the project is to initiate systematic training from earliest stages of the female scientists' careers (predocs) and to build research and career management skills and tools that are relevant to the stage of the career (postdocs and beyond). It is targeting especially scientists in natural sciences and technology. This training part of the project is organized in a form of a Summer School.

A second dimension of the ADVANCE is a Mentoring and Coaching program. Systematic mentoring is one of the key factors in supporting female scientists in their career development both in academia and private sector. Together with the partner institutions, ADVANCE has developed a program offering an opportunity for professional and personal development, details of which can be looked at in its website (ADVANCE, 2006).

In addition, another main goal of ADVANCE is to have a European dimension in the training and provide an opportunity to amplify the impact of the training and mentoring program by supporting individuals to build up similar training programs in other institutions. Visibility of such training programs will also generate increased awareness of the problem among the current decision makers and the society in general.

CONCLUSION

The fact that the women scientists are underrepresented at higher level managerial positions in the domain of science and industry needs is indisputable. It's also clear that the reasons for such a phenomenon are numerous but usually different in any particular case.

A proper gender equality policy should be designed and implemented on institutional level; exchange of good practices should be organized and lots of efforts should be invested in networking activities. All these could support young female scientists in their career development and advancement thus contributing to sustainable development and social progress in gradually integrating Europe.

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УЧЕБНИ ПЛАНОВЕ В БЪЛГАРСКОТО СРЕДНО УЧИЛИЩЕ В ПЕРИОДА IXX – XXI ВЕК

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Резюме

Разгледани са учебните програми за българското средно училище от създаването на Третото Българско царство до днес. Направен е анализ на промените в плановете за различните предмети и културно образователни области. Получените резултати позволяват да се направят някои изводи за развитието на българското средно образование.

Keywords
Средно образование, учебни планове, учебни предмети.

УВОД

Средното образование е една от основите, определящи развитието на всяка нация. С развитието на обществото се променя и развива и средното образование. То винаги трябва да отговаря на съвременното ниво на обществото, да подготвя хора, които ще живеят при определени условия и които ще могат да се адаптират при промяна на тези условия.

В момента България се намира в началото на процеси свързани с въвеждане на пазарната икономика и включването на страната в Европа. Важна страна на тези промени е свързана със средното образование. В него трябва да се извършат значителни промени, така че от една страна да може да се подготви младежта за новите условия на живот и от друга да доближи образованието до европейските стандарти.

Необходимостта от промени в Българското образование предизвика дискусии на тема какво е било то в миналото и дали действително е необходима промяна. Отговорът на втория въпрос е ясен и недвусмислен – промяна е необходима. Отговорът на първия въпрос е по-сложен и изисква известно връщане назад и коментиране на учебните планове на гимназиалния етап на обучението

Настоящата статия няма претенцията да реши въпроса как трябва да протече образователната реформа. Тя просто ще покаже някои извървени крачки и как се е стигнало до сегашното положение. Прегледът ще позволи да се потърсят някои забравени рационални страни на средното образование.

В работата, поради ограничения обем, няма да се разглеждат различните видове гимназии, а само отделни учебните планове.

РЕЗУЛТАТИ И ДИСКУСИЯ

Съвременното средно образование у нас започва да се развива преди Освобождението. При съставянето на първите учебни планове и програми за средното училище у нас се използва опитът на западното средно училище, съобразен с конкретните условия у нас - слабо развита промишленост и сравнително ниска култура на населението.

№	Учебни предмети	Класове				
		VIII	IX	X	XI	Всичко
1.	Вероучение	1	1	1	-	3
2.	Български език и литература	2	2	2	3	9
3.	Руски език	2	2	-	-	4
4.	Западноевропейски езици	4	3	3	3	13
5.	История	2	2	2	1	7
6.	Политическа икономия	-	-	-	2	2
7.	Психология	-	-	-	2	2
8.	Математика	6	4	4	4	18
9.	Геометрично чертане	2	3	3	3	11
10.	Физика	-	3	3	4	10
11.	Химия	3	2	2	1	8
12.	Естествена история	3	2	3	1	9
13.	Рисуване	4	4	4	4	16
	Всичко	**29**	**28**	**27**	**28**	**112**

Таблица 1. Учебен план за гимназиите в Княжество България от 1882г

№	Учебни предмети	Класове				
		VIII	IX	X	XI	Всичко
1.	Вероучение	-	-	2	2	4
2.	Български език и литература	5	4	3	3	15
3.	Руски език	2	2	-	-	4
4.	Западноевропейски езици	3	3	3	3	12
5.	История	2	2	3	2	9
6.	Логика	-	-	-	2	2
7.	Психология	-	-	1	-	1
8.	Математика	5	5	5	4	19
9.	Геометрично чертане	2	-	-	-	2
10.	Дескриптивна геометрия	-	3	3	3	9
11.	Физика	-	2	3	3	8
12.	Химия	2	2	2	3	9
13.	Естествена история	2	2	2	2	8
14.	Рисуване	4	4	2	4	14
15.	Пение	2	-	-	-	2
16.	Гимнастика	1	1	1	1	4
	Всичко	**30**	**30**	**31**	**30**	**121**

Таблица 2. Учебен план за гимназиите в Княжество България от 1890г

Първите учебни планове /Таблици 1-2/ съдържат почти всички учебни предмети, които се изучават и сега, включително предмети от гражданското обучение – логика и психология. Прави впечатление, че липсва географията, която е била част от предмета естествознание, заедно с биологичните знания. В първия план /Таблица 1/ прави впечатление сравнително малкото часове за български език и липсата на пеене и спорт, По всяка вероятност се е считало, че българският език е трябвало да бъде усвоен в началното училище и прогимназиалния етап. В другите учебни планове тази слабости са отстранени. Още от първия учебен план се вижда значителния дял на обучението по чужди езици – общо 17 часа, което се запазва във всички учебни планове и до сега.

№	Учебни предмети	Класове					
		VIII	IX	X	XI	XII	Всичко
1.	Вероучение	2	1	-	1	-	4
2.	Бълг. език и литература	4	4	3	3	5	19
3.	Руски език	-	2	2	-	-	4
4.	Западноевропейски езици	4	4	3	3	3	17
5.	Латински език	-	-	2	2	2	6
6.	История	2	2	1	2	3	10
7.	География	-	2	1	1	2	6
8.	Гражданско учение	-	-	-	-	1	1

9.	Философска пропедевтика	-	-	2	2	1	5
10.	Математика	4	3	4	3/5	4/3	18/8
11.	Физика	-	-	2	2	1	5
12.	Химия	3	3	2	1	1	10
13.	Естествена история	3	3	2	-	2	10
14.	Хигиена	-	-	-	2	-	2
15.	Домакинство и майчинство	-	1	-	2	1	4
16.	Гимнастика	2	2	2	2	2	10
17.	Пение	1	1	1	1	1	5
18.	Рисуване	2	2	2	2	-	8
19.	Ръчна работа и ръкоделие	2	-	-	-	-	2
20.	Стенография	/2/	-	-	-	-	/2/
	Всичко	**27/2**	**30**	**30**	**30/5**	**31/3**	**158/10**

Таблица 3. Учебен план за гимназиите в Царство България от 1940 г

Учебният план от 1940г /Таблица 3/ показва, че обучението в гимназията е нараснало на 5 години. В девическите и мъжките гимназии са изучавани различни специфични предмети свързани с бита. Значително учебно време е отделено за чуждоезиковото и математическото обучение. Появили са се и първите избираеми учебни предмети. Географията и биологията са разделени. За първи път се въвежда и здравно образование под формата на предмета хигиена.

№	**Учебни предмети**	**Класове**				
		VIII	**IX**	**X**	**XI**	**Всичко**
1.	Бълг. език и литература	4	4	4	5	17
2.	Руски език	3	3	3	3	12
3.	Западноевропейски езици	3	3	3	3	12
4.	Латински език	(2)	(2)	(2)	(2)	(8)
5.	Гръцки език	-	-	(2)	(2)	(4)
6.	История	2	2	3	3	10
7.	География	2	2	2	-	6
8.	Конституция	-	-	-	1	1
9.	Психология	-	-	-	2	2
10.	Математика	5	5	5	4	19
11.	Геометрично чертане	1	1	1	1	4
12.	Физика	2	3	3	3	11
13.	Космография (астрономия)	-	-	-	1	1
14.	Химия	3	2	2	2	9
15.	Естествени науки	1	2	2	-	5
16.	Гимнастика	2	2	2	2	8
17.	Пение	1	1	-	-	2

18.	Рисуване	1	-	-	-	1
19.	Стенография	(2)	(2)			(4)
	Всичко	**29/4**	**30/4**	**30/4**	**30/4**	**150/16**

Таблица 4. Учебен план за горен курс
– средно трудово политехническо училище от 1955г.

Учебния план за 1955 г /Таблица 4/ показва, че наименованието "гимназия" е изчезнало и средно образование се получава в така нареченото "средно трудово политехническо училище". В учебния план липсва политехнически елемент, но учениците всяка седмица участваха по половин ден в така наречената "практика", която се провеждаще в различни промишлени прадприятия, която не е отбелязана в плана. Значителен е хорариума по физика 10 часа, като е включена и астрономия. Естествените науки са само 5 часа като освен класическа биология се изучаваше е предмета дарвинизъм, в който се отричаха всички съвремени постижения на биологията. Може би това е най-неудачния учебен план както по структура така и по съдържание на отделните учебни предмети. Общият курс на обучение е намален отново на 4 години.

В учебния план от 1967 г /Таблица 5/ са запазени някои от слабите страни на плана от 1955 г., включително и наименованието гимназиално образование. Все пак към "политехническо" е прибавено "общообразователно", което е определен напредък. Към плана са прибавени елементи на трудово обучение и е засилен идеологическия елемент в обучението чрез предмета "основи на комунизма". Броя на избираемите предмети е нараснал и за първи път е въведен турския език като избираем. Бяха направени сериозни усилия за получаване на определени професии по време на средното образоваие, като бяха изградени центрове, в които се провеждаше практическото обучение на учениците и се получаваше правоспособност за някои професии като: продавач, сервитьор, шивач и др. Резултатите от тази идея не бяха добри и трудовото обучение в този вид беше отречено и премахнато.

№	**Задължителни учебни предмети**	**Класове**				
		VIII	**IX**	**X**	**XI**	**Всичко**
1.	Български език и литература	5	4	3	4	16
2.	Руски език	2/2/	2/2/	1/2/	1/2/	10/8/
3.	Западноевропейски езици	2/2/	3/2/	3/2/	2/2/	10/8/
4.	История и конституция	2	2	2	2	8
5.	География	2	2	2	-	6
6.	Психология и логика	-	-	2	-	2
7.	Основи на комунизма	-	-	-	2	2
8.	Математика	4	5	4	5	18
9.	Физика и астрономия	2	3	3	4	12
10.	Химия	2	2	3	2	9
11.	Биология	2	2	2	2	8
12.	Физически възпитание	2	2	2	2	8
13.	Музика	1	1/2/	/2/	/2/	2/6/
14.	Рисуване	/2/	/2/	/2/	/2/	/8/
	Всичко	**26/6/**	**28/8/**	**27/8/**	**24/8/**	**105/30/**
	Трудово обучение					

1.	Основи на селската стопанство	2	-	-	-	2
2.	Основи на промишлеността	2	3	-	1	6
3.	Практика и теория на произв. обучение			3	5	8
	Всичко	**4**	**3**	**3**	**6**	**16**
	Избираеми предмети					
1.	Стенография	2	2	-	-	4
2.	Латински език	-	2	2	-	4
3.	Турски език	2	2	2	2	8
4.	Мотокар и трактор	-	2	2	-	4
	Всичко	**4**	**8**	**6**	**2**	**20**

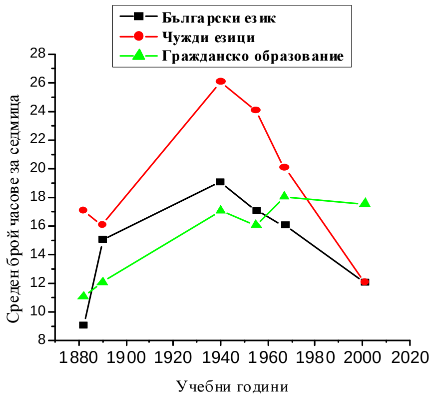
Таблица 5. Учебен план за
общообразователно политехническо училище от 1967г.

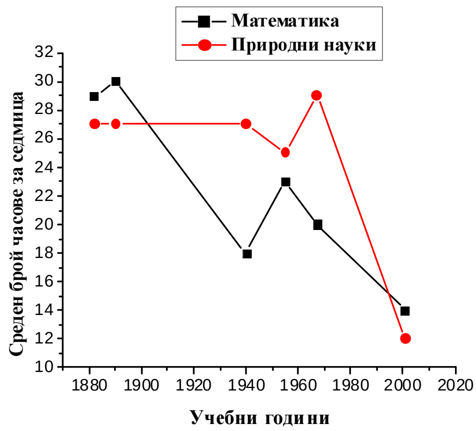
Учебният план от 2000 г /Таблица 5/ може да се разгледа като учебен план, в който са събрани положителния опит от учебните планове използвани досега, като са взети предвид и съвременните изисквания за средно образование. Приета е и нова философия в обучението като се дава възможност за елементи на индивидуален подход в обучението. Превидени са значителен брой задължително и свободно избираеми часове. Въведени са значителен брой нови предмети, някои от които са съществували и в старите учебни планове. Прави впечатление, че гимназиалното образование започва не от VIII, а от IX клас. Силно е намалено задължителното обучение по природните дисциплини. Значително са нараснали часовете за спорт. Въпреки интересните идеи този учебен план се нуждае от подобрение, поради което промените в него продължават ежегодно, като разбира се са не винаги удачни. Все пак основата, която е положена е в правилна посока.

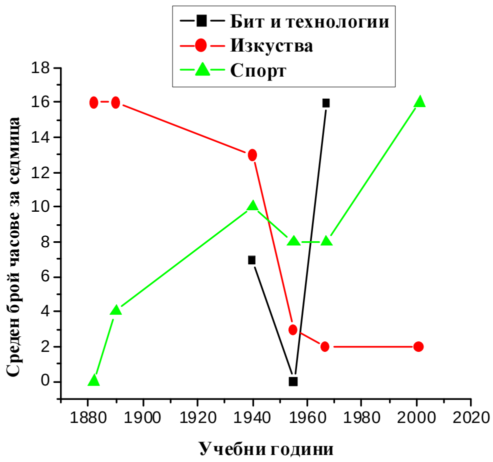
№	Културно образователни области – частично уедрени	Класове				
		IX	X	XI	XII	Общо
1.	**Български език и литература и чужди езици**					**24**
	Български език и литература	3	3	3	3	12
	Чужд език I	2	2	2	2	8
	Чужд език II	2	2	-	-	4
2.	**Математика, информатика и информ. технологии**					**14**
	Математика	3	3	2	2	10
	Информатика	2	-	-	-	2
	Информационни технологии	1	1	-	-	2
3.	**Обществени науки и гражданско образование**					**17,5**
	История и цивилизация	2	2	2	-	6
	География и икономика	1,5	1,5	1	-	4
	Психология и логика	1,5	-	-	-	1,5
	Етика и право	-	1,5	-	-	1,5
	Философия	-	-	1,5	-	1,5
	Свят и личност	-	-	-	2	3
4.	**Природни науки и екология**					**12**

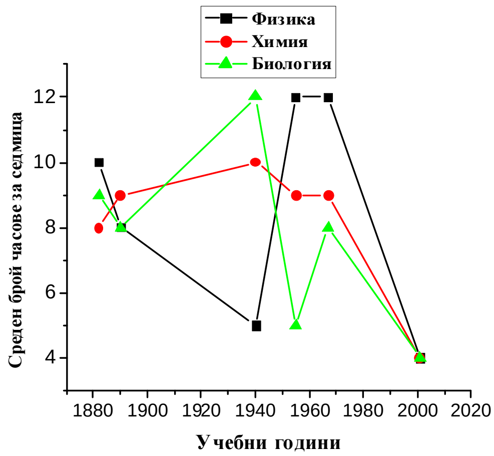
	Биология и здравно образование	2	2	-	-	4
	Физика и астрономия	2	2	-	-	4
	Химия и опазване на околната среда	2	2	-	-	4
5.	**Изкуства**					**2**
	Музика	1	-	-	-	1
	Изобразително изкуство	1	-	-	-	1
6.	**Физическа култура и спорт**					**16**
	Физическа култура и спорт	4	4	4	4	16
	Всичко	30	26	15,5	9	80,5
7.	**Задължително избираема подготовка**	4	8	17,5	21	50,5
8.	**Свободно избираема подготовка**	4	4	4	4	16
	Всичко избираема подготовка	**8**	**12**	**21,5**	**25**	**66,5**
	Всичко: ЗП + ЗИП + СИП	**38**	**38**	**37**	**34**	**147**

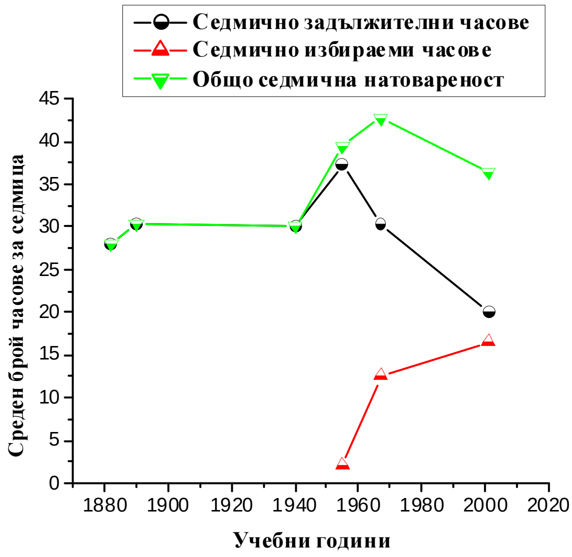
Таблица 6. Учебен план за гимназии след завършено основно образование от 2000 г.











Представлява интерес сравняването на хорариума на различните учебни предмети през годините. От представените фигури се вижда, че изучаването на българския език и чуждите езици е най-интензивно през 1940 г., като след това намалява. Интересна е тенденцията на гражданското образование, хорариума на което (реализиран в различни предмети през различните години) нараства плавно и достига до насищане. Наблюдава се намаление на часовете предвидени за математика и природните науки. Особенно недобро впечатление прави значителното нарастване на часовете за спорт и намаляването на часовете определени за изучаване на изкуствата. Предметите от „Бит и технологии” се появяват и изчезват.

Особено интересно е значителното вариране на часовете определени за различните природни предмети – особено при биологията и физиката. Направения преглед показа, че в известна степен това е свързано с личностите, които са били министри на образованието, както и поради значението на една или друга наука за определен период.

Както може и да се очаква общата натовареност на ученици нараства през годините, като все пак се наблюдава известно облекчаване при последния учебен план. Както трябва да се очаква хорариума за избираемите предмети нараства, а този на задължителните намалява. От последната фигура се вижда, че часовете за задължителна подготовка и избираемите часове са приблизително равни, което е в съзвучие с тенденциите в съвременното средно образование.

ЗАКЛЮЧЕНИЕ

Настоящата работа позволява да се направят следните изводи:

1. Българското средно образование е било модерно светско образование от края на IXX век до сега.

2. И днес старите учебните планове звучат съвременно. В тях има елементи, които сега се смятат за новост– избираеми предмети, гражданско образование, предмети с практическа насоченост, засилено езиково обучение.

3. Наблюдава се вариране в учебната натовареност при природните дисциплини, което според нас е свързано с определени личностни и политически причини.

4. Прави впечатление значителното намаляване на задължителните часове за обучение по природните науки и математиката и чуждите езици в сегашния учебен план в сравнение със старите учебни планове.

5. Считаме, че традициите в българското средно училище по отношение на учебните планове са основа за нормално провеждане на реформата.

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