## Contents

Modelling Real World Using Stochastic Processes and Filtration
By Peter Jaeger ..... 1
Circumcenter, Circumcircle and Centroid of a Triangle By Roland Coghetto ..... 17
Altitude, Orthocenter of a Triangle and Triangulation By Roland Coghetto ..... 27
Divisible $\mathbb{Z}$-modules
By Yuichi Futa and Yasunari Shidama ..... 37
Lattice of $\mathbb{Z}$-module
By Yuichi Futa and Yasunari Shidama ..... 49
Product Pre-Measure
By Noboru Endou ..... 69
Conservation Rules of Direct Sum Decomposition of Groups
By Kazuhisa Nakasho et al. ..... 81

# Modelling Real World Using Stochastic Processes and Filtration 

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#### Abstract

Summary. First we give an implementation in Mizar [2] basic important definitions of stochastic finance, i.e. filtration ([9], pp. 183 and 185), adapted stochastic process ( 9 , p. 185) and predictable stochastic process (6, p. 224). Second we give some concrete formalization and verification to real world examples.

In article [8] we started to define random variables for a similar presentation to the book [6]. Here we continue this study. Next we define the stochastic process. For further definitions based on stochastic process we implement the definition of filtration.

To get a better understanding we give a real world example and connect the statements to the theorems. Other similar examples are given in [10], pp. 143159 and in [12], pp. 110-124. First we introduce sets which give informations referring to today ( $\Omega_{\text {now }}$, Def.6), tomorrow $\left(\Omega_{f u t_{1}}\right.$, Def.7) and the day after tomorrow ( $\Omega_{f u t_{2}}$, Def.8). We give an overview for some events in the $\sigma$-algebras $\Omega_{\text {now }}, \Omega_{\text {fut } 1}$ and $\Omega_{f u t 2}$, see theorems (22) and (23).

The given events are necessary for creating our next functions. The implementations take the form of: $\Omega_{n o w} \subset \Omega_{f u t 1} \subset \Omega_{f u t 2}$ see theorem (24). This tells us growing informations from now to the future $1=$ now, $2=$ tomorrow, $3=$ the day after tomorrow.

We install functions $f:\{1,2,3,4\} \rightarrow \mathbb{R}$ as following: $f_{1}: x \rightarrow 100, \forall x \in \operatorname{dom} f$, see theorem (36), $f_{2}: x \rightarrow 80$, for $x=1$ or $x=2$ and $f_{2}: x \rightarrow 120$, for $x=3$ or $x=4$, see theorem (37), $f_{3}: x \rightarrow 60$, for $x=1, f_{3}: x \rightarrow 80$, for $x=2$ and $f_{3}: x \rightarrow 100$, for $x=3, f_{3}: x \rightarrow 120$, for $x=4$ see theorem (38). These functions are real random variable: $f_{1}$ over $\Omega_{n o w}, f_{2}$ over $\Omega_{f u t 1}, f_{3}$ over $\Omega_{f u t 2}$, see theorems (46), (43) and (40). We can prove that these functions can be used for giving an example for an adapted stochastic process. See theorem


 (49).We want to give an interpretation to these functions: suppose you have an equity $A$ which has now $\left(=w_{1}\right)$ the value 100 . Tomorrow $A$ changes depending which scenario occurs - e.g. another marketing strategy. In scenario 1 ( $=w_{11}$ ) it has the value 80 , in scenario $2\left(=w_{12}\right)$ it has the value 120 . The day after tomorrow $A$ changes again. In scenario $1\left(=w_{111}\right)$ it has the value 60 , in scenario $2\left(=w_{112}\right)$ the value 80 , in scenario $3\left(=w_{121}\right)$ the value 100 and in scenario 4 ( $=w_{122}$ ) it has the value 120 . For a visualization refer to the tree:

$$
\begin{aligned}
& \text { Now tomorrow the day after tomorrow } \\
& w_{111}=\{1\} \\
& w_{11}=\{1,2\} \quad< \\
& w_{112}=\{2\} \\
& w_{1}=\{1,2,3,4\} \quad< \\
& w_{12}=\{3,4\}<\begin{array}{l}
w_{121}=\{3\} \\
w_{122}=\{4\}
\end{array}
\end{aligned}
$$

The sets $w_{1}, w_{11}, w_{12}, w_{111}, w_{112}, w_{121}, w_{122}$ which are subsets of $\{1,2,3,4\}$, see (22), tell us which market scenario occurs. The functions tell us the values to the relevant market scenario:

| Now | tomorrow |  | the day after tomorrow $f_{3}\left(w_{i}\right)=60, w_{i}$ in $w_{111}$ |
| :---: | :---: | :---: | :---: |
|  | $\begin{aligned} & f_{2}\left(w_{i}\right)=80 \\ & w_{i} \text { in } w_{11} \end{aligned}$ | $<$ | $f_{3}\left(w_{i}\right)=80, w_{i} \text { in } w_{112}$ |
| $\begin{aligned} & f_{1}\left(w_{i}\right)=100 \\ & w_{i} \text { in } w_{1} \end{aligned}$ |  |  | $f_{3}\left(w_{i}\right)=100, w_{i}$ in $w_{121}$ |
|  | $f_{2}\left(w_{i}\right)=120$ | $<$ |  |
|  | $w_{i}$ in $w_{12}$ |  | $f_{3}\left(w_{i}\right)=120, w_{i}$ in $w_{122}$ |

For a better understanding of the definition of the random variable and the relation to the functions refer to [7], p. 20. For the proof of certain sets as $\sigma$-fields refer to [7, pp. 10-11 and [9], pp. 1-2.

This article is the next step to the arbitrage opportunity. If you use for example a simple probability measure, refer, for example to literature [3], pp. 28-34, [6], p. 6 and p. 232 you can calculate whether an arbitrage exists or not. Note, that the example given in literature [3] needs 8 instead of 4 informations as in our model. If we want to code the first 3 given time points into our model we would have the following graph, see theorems (47), (44) and (41):


The function for the "Call-Option" is given in literature [3], p. 28. The function is realized in Def.5. As a background, more examples for using the definition of filtration are given in 9, pp. 185-188.

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## 1. Preliminaries

Now we state the proposition:
(1) Let us consider objects $a, b$. If $a \neq b$, then $\{a\} \subset\{a, b\}$.

Let $I$ be a non empty subset of $\mathbb{N}$. Observe that $I\left(\in 2^{\mathbb{R}}\right)$ is non empty.
Let us consider an element $T$ of $\mathbb{N}$. Now we state the propositions:
(2) $\{w$, where $w$ is an element of $\mathbb{N}: w>0$ and $w \leqslant T\} \subseteq\{w$, where $w$ is an element of $\mathbb{N}: w \leqslant T\}$.
(3) $\{w$, where $w$ is an element of $\mathbb{N}: w \leqslant T\}$ is a non empty subset of $\mathbb{N}$.
(4) If $T>0$, then $\{w$, where $w$ is an element of $\mathbb{N}: w>0$ and $w \leqslant T\}$ is a non empty subset of $\mathbb{N}$.
Proof: $\{w$, where $w$ is an element of $\mathbb{N}: w>0$ and $w \leqslant T\}$ is a subset of $\mathbb{N} .1>0$ and $1 \leqslant T$ by [1, (24)].
Now we state the proposition:
(5) Let us consider a non empty set $\Omega$. Then $\Omega \longmapsto 1$ is a function from $\Omega$ into $\mathbb{R}$.

## 2. Special Random Variables

Now we state the proposition:
(6) Let us consider a natural number $d$, a sequence $\varphi$ of real numbers, a non empty set $\Omega$, a $\sigma$-field $F$ of subsets of $\Omega$, a non empty set $X$, a sequence $G$ of $X$, and an element $w$ of $\Omega$. Then $\{$ the portfolio value for future extension of $d, \varphi, F, G$ and $w\}$ is an event of the Borel sets.
Let $d$ be a natural number, $\varphi$ be a sequence of real numbers, $\Omega$ be a non empty set, $F$ be a $\sigma$-field of subsets of $\Omega, X$ be a non empty set, $G$ be a sequence of $X$, and $w$ be an element of $\Omega$. Note that the portfolio value for future extension of $d, \varphi, F, G$ and $w$ yields an element of $\mathbb{R}$. The $\mathcal{R} \mathcal{V}$-portfolio value for future extension of $\varphi, F, G$ and $d$ yielding a function from $\Omega$ into $\mathbb{R}$ is defined by
(Def. 1) for every element $w$ of $\Omega, i t(w)=$ the portfolio value for future extension of $d, \varphi, F, G$ and $w$.

Let us observe that the $\mathcal{R} \mathcal{V}$-portfolio value for future extension of $\varphi, F, G$ and $d$ yields a random variable of $F$ and the Borel sets. Let $w$ be an element of $\Omega$. Let us note that the portfolio value for future of $d, \varphi, F, G$ and $w$ yields a real number and is defined by the term
(Def. 2) $\quad\left(\sum_{\alpha=0}^{\kappa}((\right.$ the elements of the random variables for the future elements of portfolio value of $(\varphi, F, G, w)) \uparrow 1)(\alpha))_{\kappa \in \mathbb{N}}(d-1)$.
Let us note that the portfolio value for future of $d, \varphi, F, G$ and $w$ yields an element of $\mathbb{R}$. The $\mathcal{R} \mathcal{V}$-portfolio value for future of $\varphi, F, G$ and $d$ yielding a function from $\Omega$ into $\mathbb{R}$ is defined by
(Def. 3) for every element $w$ of $\Omega, i t(w)=$ the portfolio value for future of $d+1$, $\varphi, F, G$ and $w$.
Let us note that the $\mathcal{R} \mathcal{V}$-portfolio value for future of $\varphi, F, G$ and $d$ yields a random variable of $F$ and the Borel sets. Now we state the propositions:
(7) Let us consider a natural number $d$, a sequence $\varphi$ of real numbers, a non empty set $\Omega$, a $\sigma$-field $F$ of subsets of $\Omega$, a non empty set $X$, a sequence $G$ of $X$, and an element $w$ of $\Omega$. Then
(i) the portfolio value for future of $d+1, \varphi, F, G$ and $w=($ the $\mathcal{R} \mathcal{V}$-portfolio value for future of $\varphi, F, G$ and $d)(w)$, and
(ii) $\{$ the portfolio value for future of $d+1, \varphi, F, G$ and $w\}$ is an event of the Borel sets.
(8) Let us consider a non empty set $\Omega$, a $\sigma$-field $F$ of subsets of $\Omega$, a non empty set $X$, a sequence $G$ of $X$, a sequence $\varphi$ of real numbers, and a natural number $d$. Then the $\mathcal{R} \mathcal{V}$-portfolio value for future extension of $\varphi, F, G$ and $d+1=($ the $\mathcal{R} \mathcal{V}$-portfolio value for future of $\varphi, F, G$ and $d)+$ (the random variables for the future elements of portfolio value of $(\varphi, F, G, 0))$.
(9) Let us consider non empty sets $\Omega, \Omega_{2}$, a $\sigma$-field $\Sigma$ of subsets of $\Omega$, a $\sigma$ field $\Sigma_{2}$ of subsets of $\Omega_{2}$, and an element $s$ of $\Omega_{2}$. Then $\Omega \longmapsto s$ is random variable on $\Sigma$ and $\Sigma_{2}$.
(10) Let us consider a non empty set $\Omega$, a $\sigma$-field $\Sigma$ of subsets of $\Omega$, a random variable $\mathcal{R} \mathcal{V}$ of $\Sigma$ and the Borel sets, and an element $K$ of $\mathbb{R}$. Then $\mathcal{R} \mathcal{V}$ $(\Omega \longmapsto K)$ is a random variable of $\Sigma$ and the Borel sets. The theorem is a consequence of (9).
Let $\Omega$ be a non empty set, $\mathcal{R} \mathcal{V}$ be a function from $\Omega$ into $\mathbb{R}$, and $w$ be an element of $\Omega$. The functor Set-Call-Option $(\mathcal{R} \mathcal{V}, w)$ yielding an element of $\mathbb{R}$ is defined by the term
(Def. 4) $\begin{cases}\mathcal{R} \mathcal{V}(w), & \text { if } \mathcal{R} \mathcal{V}(w) \geqslant 0, \\ 0, & \text { otherwise. }\end{cases}$

Let $\Sigma$ be a $\sigma$-field of subsets of $\Omega, \mathcal{R} \mathcal{V}$ be a random variable of $\Sigma$ and the Borel sets, and $K$ be an element of $\mathbb{R}$. The Call-Option on $\mathcal{R} \mathcal{V}$ and $K$ yielding a function from $\Omega$ into $\mathbb{R}$ is defined by
(Def. 5) for every element $w$ of $\Omega$, if $(\mathcal{R} \mathcal{V}-(\Omega \longmapsto K))(w) \geqslant 0$, then $i t(w)=$ $(\mathcal{R} \mathcal{V}-(\Omega \longmapsto K))(w)$ and if $(\mathcal{R} \mathcal{V}-(\Omega \longmapsto K))(w)<0$, then $i t(w)=0$.

## 3. Special $\sigma$-Fields

Let us consider a sequence $A_{1}$ of subsets of $\{1,2,3,4\}$ and a real number $w$. Now we state the propositions:
(11) Suppose $w=1$ or $w=3$. Then suppose for every natural number $n$, $A_{1}(n)=\emptyset$ or $A_{1}(n)=\{1,2\}$ or $A_{1}(n)=\{3,4\}$ or $A_{1}(n)=\{1,2,3,4\}$. Then $\{w\} \neq$ Intersection $A_{1}$.
(12) Suppose $w=2$ or $w=4$. Then suppose for every natural number $n$, $A_{1}(n)=\emptyset$ or $A_{1}(n)=\{1,2\}$ or $A_{1}(n)=\{3,4\}$ or $A_{1}(n)=\{1,2,3,4\}$. Then $\{w\} \neq$ Intersection $A_{1}$.
Now we state the propositions:
(13) Let us consider sets $M, A_{1}, A_{2}$. Suppose $M=\{\emptyset,\{1,2,3,4\}\}$ and $A_{1}$, $A_{2} \in M$. Then $A_{1} \cap A_{2} \in M$.
(14) Let us consider a sequence $A_{1}$ of subsets of $\{1,2,3,4\}$. Suppose for every natural number $n$ and for every natural number $k, A_{1}(n) \cap A_{1}(k) \neq \emptyset$ and $\operatorname{rng} A_{1} \subseteq\{\emptyset,\{1,2\},\{3,4\},\{1,2,3,4\}\}$. Then
(i) Intersection $A_{1}=\emptyset$, or
(ii) Intersection $A_{1}=\{1,2\}$, or
(iii) Intersection $A_{1}=\{3,4\}$, or
(iv) Intersection $A_{1}=\{1,2,3,4\}$.

Proof: For every natural number $n, A_{1}(n) \in\{\emptyset,\{1,2\},\{3,4\},\{1,2,3,4\}\}$ by [1, (20)], [4, (3)]. For every natural number $n, A_{1}(n)=\emptyset$ or $A_{1}(n)=$ $\{1,2\}$ or $A_{1}(n)=\{3,4\}$ or $A_{1}(n)=\{1,2,3,4\}$.
Let us consider a sequence $A_{1}$ of subsets of $\{1,2,3,4\}$ and a set $M$. Now we state the propositions:
(15) Suppose $M=\{\emptyset,\{1,2\},\{3,4\},\{1,2,3,4\}\}$ and Intersection $A_{1}=\{1,2,3,4\}$. Then Intersection $A_{1} \in M$.
(16) Suppose $M=\{\emptyset,\{1,2\},\{3,4\},\{1,2,3,4\}\}$ and Intersection $A_{1}=\{3,4\}$. Then Intersection $A_{1} \in M$.
(17) Suppose $M=\{\emptyset,\{1,2\},\{3,4\},\{1,2,3,4\}\}$ and Intersection $A_{1}=\{1,2\}$. Then Intersection $A_{1} \in M$.
(18) Suppose $M=\{\emptyset,\{1,2\},\{3,4\},\{1,2,3,4\}\}$ and Intersection $A_{1}=\emptyset$. Then Intersection $A_{1} \in M$.
Now we state the propositions:
(19) Let us consider a set $M$, and a sequence $A_{1}$ of subsets of $\{1,2,3,4\}$. Suppose rng $A_{1} \subseteq M$ and $M=\{\emptyset,\{1,2\},\{3,4\},\{1,2,3,4\}\}$.
Then Intersection $A_{1} \in M$.
Proof: Intersection $A_{1} \in\{\emptyset,\{1,2\},\{3,4\},\{1,2,3,4\}\}$ by [11, (13)], (14).
(20) Let us consider sets $M, M_{1}$, and a sequence $A_{1}$ of subsets of $M_{1}$. Suppose $M_{1}=\{1,2,3,4\}$ and $\operatorname{rng} A_{1} \subseteq M$ and $M=\{\emptyset,\{1,2,3,4\}\}$. If Intersection $A_{1} \neq \emptyset$, then Intersection $A_{1} \in M$.
Proof: For every natural number $n, A_{1}(n)=\emptyset$ or $A_{1}(n)=\{1,2,3,4\}$ by [1, (20)], [4, (3)]. If there exists a natural number $n$ such that $A_{1}(n)=\emptyset$, then Intersection $A_{1}=\emptyset$ by [11, (13)]. Intersection $A_{1}=\{1,2,3,4\}$ by [11, (13)].
(21) Let us consider sets $M, M_{1}$, and a sequence $A_{1}$ of subsets of $M_{1}$. Suppose $M_{1}=\{1,2,3,4\}$ and $\mathrm{rng} A_{1} \subseteq M$ and $M=\{\emptyset,\{1,2,3,4\}\}$. Let us consider a natural number $k_{1}$, and a natural number $k_{2}$. Then $A_{1}\left(k_{1}\right) \cap A_{1}\left(k_{2}\right) \in M$. Proof: $k_{1} \in \operatorname{dom} A_{1}$ by [1, (20)]. $k_{2} \in \operatorname{dom} A_{1}$ by [1, (20)]. $A_{1}\left(k_{1}\right) \cap$ $A_{1}\left(k_{2}\right) \in M$.
The functor $\Omega_{\text {now }}$ yielding a $\sigma$-field of subsets of $\{1,2,3,4\}$ is defined by the term
(Def. 6) $\{\emptyset,\{1,2,3,4\}\}$.
The functor $\Omega_{f u t 1}$ yielding a $\sigma$-field of subsets of $\{1,2,3,4\}$ is defined by the term
(Def. 7) $\{\emptyset,\{1,2\},\{3,4\},\{1,2,3,4\}\}$.
The functor $\Omega_{f u t 2}$ yielding a $\sigma$-field of subsets of $\{1,2,3,4\}$ is defined by the term
(Def. 8) $2^{\{1,2,3,4\}}$.
Let us consider a set $\Omega$.
Let us assume that $\Omega=\{1,2,3,4\}$. Now we state the propositions:
(22) (i) $\{1\} \subseteq \Omega$, and
(ii) $\{2\} \subseteq \Omega$, and
(iii) $\{3\} \subseteq \Omega$, and
(iv) $\{4\} \subseteq \Omega$, and
(v) $\{1,2\} \subseteq \Omega$, and
(vi) $\{3,4\} \subseteq \Omega$, and
(vii) $\emptyset \subseteq \Omega \subseteq \Omega$.
(23) (i) $\Omega, \emptyset \in \Omega_{\text {now }}$, and
(ii) $\{1,2\},\{3,4\}, \Omega, \emptyset \in \Omega_{f u t 1}$, and
(iii) $\Omega, \emptyset,\{1\},\{2\},\{3\},\{4\} \in \Omega_{\text {fut } 2}$.

Now we state the proposition:
$\Omega_{\text {now }} \subset \Omega_{f u t 1} \subset \Omega_{f u t 2}$.

## 4. Construction of Filtration and Examples

Now we state the propositions:
(25) There exists a non empty set $\Omega$ and there exist $\sigma$-fields $F_{1}, F_{2}, F_{3}$ of subsets of $\Omega$ such that $F_{1} \subset F_{2} \subset F_{3}$.
(26) There exist non empty sets $\Omega_{1}, \Omega_{2}, \Omega_{3}, \Omega_{4}$ such that
(i) $\Omega_{1} \subset \Omega_{2} \subset \Omega_{3} \subset \Omega_{4}$, and
(ii) there exists a $\sigma$-field $F_{1}$ of subsets of $\Omega_{1}$ and there exists a $\sigma$-field $F_{2}$ of subsets of $\Omega_{2}$ and there exists a $\sigma$-field $F_{3}$ of subsets of $\Omega_{3}$ and there exists a $\sigma$-field $F_{4}$ of subsets of $\Omega_{4}$ such that $F_{1} \subseteq F_{2} \subseteq F_{3} \subseteq F_{4}$.

Let $I, \Omega$ be non empty sets, $\Sigma$ be a $\sigma$-field of subsets of $\Omega, M$ be a many sorted $\sigma$-field over $I$ and $\Sigma$, and $i$ be an element of $I$. The functor $\mathcal{M}_{\sigma \text {-field }}(M, i)$ yielding a $\sigma$-field of subsets of $\Omega$ is defined by the term
(Def. 9) $\quad M(i)$.
Let $\Omega$ be a non empty set and $I$ be a non empty subset of $\mathbb{R}$.
A filtration of $I$ and $\Sigma$ is a many sorted $\sigma$-field over $I$ and $\Sigma$ and is defined by
(Def. 10) for every elements $s, t$ of $I$ such that $s \leqslant t$ holds $i t(s)$ is a subset of $i t(t)$ and for every element $t$ of $I, i t(t) \subseteq \Sigma$.
Let $F$ be a filtration of $I$ and $\Sigma$ and $i$ be an element of $I$. The $i-\mathcal{E} \mathcal{F}$ of $F$ yielding a $\sigma$-field of subsets of $\Omega$ is defined by the term
(Def. 11) $F(i)$.
Let $k$ be an element of $\{1,2,3\}$. The functor $\operatorname{Select12-~} \sigma$-field $(k)$ yielding a subset of $2^{\{1,2,3,4\}}$ is defined by the term
(Def. 12)

$$
\begin{cases}\Omega_{n o w}, & \text { if } k=1 \\ \Omega_{f u t 1}, & \text { otherwise. }\end{cases}
$$

The functor Select123- $\sigma$-field $(k)$ yielding a subset of $2^{\{1,2,3,4\}}$ is defined by the term
(Def. 13) $\begin{cases}\operatorname{Select} 12-\sigma-\text { field }(k), & \text { if } k \leqslant 2, \\ \Omega_{f u t 2}, & \text { otherwise. }\end{cases}$
Now we state the propositions:
(27) Let us consider a $\sigma$-field $\Sigma$ of subsets of $\{1,2,3,4\}$, and a set $I$. Suppose $I=\{1,2,3\}$ and $\Sigma=2^{\{1,2,3,4\}}$. Then there exists a many sorted $\sigma$-field $M$ over $I$ and $\Sigma$ such that
(i) $M(1)=\Omega_{\text {now }}$, and
(ii) $M(2)=\Omega_{f u t 1}$, and
(iii) $M(3)=\Omega_{f u t 2}$.

Proof: Define $\mathcal{U}$ (element of $\{1,2,3\})=\operatorname{Select123-~} \sigma$-field $\left(\$_{1}\right)$. Consider $f_{4}$ being a function from $\{1,2,3\}$ into $2^{2^{\{1,2,3,4\}}}$ such that for every element $d$ of $\{1,2,3\}, f_{4}(d)=\mathcal{U}(d)$ from [5, Sch. 4]. For every set $i$ such that $i \in I$ holds $f_{4}(i)$ is a $\sigma$-field of subsets of $\{1,2,3,4\}$.
(28) Let us consider a non empty set $\Omega$, a $\sigma$-field $\Sigma$ of subsets of $\Omega$, and a non empty subset $I$ of $\mathbb{R}$. Suppose $I=\{1,2,3\}$ and $\Sigma=2^{\{1,2,3,4\}}$ and $\Omega=\{1,2,3,4\}$. Then there exists a many sorted $\sigma$-field $M$ over $I$ and $\Sigma$ such that
(i) $M(1)=\Omega_{\text {now }}$, and
(ii) $M(2)=\Omega_{f u t 1}$, and
(iii) $M(3)=\Omega_{f u t 2}$, and
(iv) $M$ is a filtration of $I$ and $\Sigma$.

The theorem is a consequence of (27).
(29) Let us consider a non empty set $\Omega$, a $\sigma$-field $\Sigma$ of subsets of $\Omega$, and a $\sigma$-field $\Sigma_{2}$ of subsets of $\{1\}$. Suppose $\Omega=\{1,2,3,4\}$. Then there exists a function $X_{1}$ from $\Omega$ into $\{1\}$ such that $X_{1}$ is random variable of $\Omega_{\text {now }}$ and $\Sigma_{2}$, random variable of $\Omega_{f u t 1}$ and $\Sigma_{2}$, and random variable of $\Omega_{f u t 2}$ and $\Sigma_{2}$.
(30) Let us consider a non empty set $\Omega$, a $\sigma$-field $\Sigma$ of subsets of $\Omega$, and a non empty subset $I$ of $\mathbb{R}$. Suppose $I=\{1,2,3\}$ and $\Sigma=2^{\{1,2,3,4\}}$ and $\Omega=\{1,2,3,4\}$. Then there exists a many sorted $\sigma$-field $M$ over $I$ and $\Sigma$ such that
(i) $M(1)=\Omega_{\text {now }}$, and
(ii) $M(2)=\Omega_{f u t 1}$, and
(iii) $M(3)=\Omega_{f u t 2}$, and
(iv) $M$ is a filtration of $I$ and $\Sigma$.

The theorem is a consequence of (27).
(31) There exist non empty sets $\Omega, \Omega_{2}$ and there exists a $\sigma$-field $\Sigma$ of subsets of $\Omega$ and there exists a $\sigma$-field $\Sigma_{2}$ of subsets of $\Omega_{2}$ and there exists a non empty subset $I$ of $\mathbb{R}$ and there exists a many sorted $\sigma$-field $Q$ over $I$ and $\Sigma$ such that $Q$ is a filtration of $I$ and $\Sigma$ and there exists a function $\mathcal{R} \mathcal{V}$ from $\Omega$ into $\Omega_{2}$ such that for every element $i$ of $I, \mathcal{R} \mathcal{V}$ is a random variable of $\mathcal{M}_{\sigma \text {-field }}(Q, i)$ and $\Sigma_{2}$. The theorem is a consequence of (30) and (29).
(32) Let us consider non empty sets $\Omega, \Omega_{2}$, a $\sigma$-field $\Sigma$ of subsets of $\Omega$, a $\sigma$ field $\Sigma_{2}$ of subsets of $\Omega_{2}$, a non empty subset $I$ of $\mathbb{R}$, and a filtration $Q$ of $I$ and $\Sigma$. Then there exists a function $\mathcal{R} \mathcal{V}$ from $\Omega$ into $\Omega_{2}$ such that for every element $i$ of $I, \mathcal{R} \mathcal{V}$ is a random variable of $\mathcal{M}_{\sigma \text {-field }}(Q, i)$ and $\Sigma_{2}$. Proof: Consider $w$ being an object such that $w \in \Omega_{2}$. Set $m_{1}=w$. Consider $m$ being a function from $\Omega$ into $\Omega_{2}$ such that $m=\Omega \longmapsto m_{1}$. For every element $i$ of $I, m$ is a random variable of $\mathcal{M}_{\sigma \text {-field }}(Q, i)$ and $\Sigma_{2}$ by [13, (7)], [11, (5), (4)].

## 5. Stochastic Process: Adapted and Predictable

Now we state the proposition:
(33) Let us consider a non empty set $\Omega$, a $\sigma$-field $\Sigma$ of subsets of $\Omega$, and a $\sigma$ field $\Sigma_{2}$ of subsets of $\Omega$. If $\Sigma_{2} \subseteq \Sigma$, then every event of $\Sigma_{2}$ is an event of $\Sigma$.
Let $\Omega, \Omega_{2}$ be non empty sets, $\Sigma$ be a $\sigma$-field of subsets of $\Omega, \Sigma_{2}$ be a $\sigma$-field of subsets of $\Omega_{2}, I$ be a non empty subset of $\mathbb{R}$, and $P$ be a probability on $\Sigma$.

A stochastic process of $I, \Sigma, \Sigma_{2}$ and $P$ is a function from $I$ into the set of random variables on $\Sigma$ and $\Sigma_{2}$ and is defined by
(Def. 14) for every element $k$ of $I$, there exists a function $\mathcal{R} \mathcal{V}$ from $\Omega$ into $\Omega_{2}$ such that $i t(k)=\mathcal{R} \mathcal{V}$ and $\mathcal{R} \mathcal{V}$ is random variable on $\Sigma$ and $\Sigma_{2}$.
Let $S$ be a stochastic process of $I, \Sigma, \Sigma_{2}$ and $P$ and $k$ be an element of $I$. The $k-\mathcal{R} \mathcal{V}$ of $S$ yielding a random variable of $\Sigma$ and $\Sigma_{2}$ is defined by the term (Def. 15) $S(k)$.

An adapted stochastic process of $I, \Sigma, \Sigma_{2}, P$ and $S$ is a function from $I$ into the set of random variables on $\Sigma$ and $\Sigma_{2}$ and is defined by
(Def. 16) there exists a filtration $k$ of $I$ and $\Sigma$ such that for every element $i$ of $I$, the $i-\mathcal{R V}$ of $S$ is random variable on the $i-\mathcal{E} \mathcal{F}$ of $k$ and $\Sigma_{2}$.

Let $I$ be a non empty subset of $\mathbb{N}, J$ be a non empty subset of $\mathbb{N}$, and $S$ be a stochastic process of $J\left(\in 2^{\mathbb{R}}\right), \Sigma, \Sigma_{2}$ and $P$.

A predictable stochastic process of $I, J, \Sigma, \Sigma_{2}, P$ and $S$ is a function from $J$ into the set of random variables on $\Sigma$ and $\Sigma_{2}$ and is defined by
(Def. 17) there exists a filtration $k$ of $I\left(\in 2^{\mathbb{R}}\right)$ and $\Sigma$ such that for every element $j$ of $J\left(\in 2^{\mathbb{R}}\right)$ for every element $i$ of $I\left(\in 2^{\mathbb{R}}\right)$ such that $j-1=i$ holds the $j-\mathcal{R} \mathcal{V}$ of $S$ is random variable on the $i-\mathcal{E} \mathcal{F}$ of $k$ and $\Sigma_{2}$.
Let $I$ be a non empty subset of $\mathbb{R}, M$ be a filtration of $I$ and $\Sigma$, and $S$ be a stochastic process of $I, \Sigma, \Sigma_{2}$ and $P$. We say that $S$ is $M$-stochastic process w.r.t. filtration if and only if
(Def. 18) for every element $i$ of $I$, the $i-\mathcal{R} \mathcal{V}$ of $S$ is random variable on the $i-\mathcal{E} \mathcal{F}$ of $M$ and $\Sigma_{2}$.

Now we state the proposition:
(34) Let us consider non empty sets $\Omega, \Omega_{2}$, a $\sigma$-field $\Sigma$ of subsets of $\Omega$, a $\sigma$ field $\Sigma_{2}$ of subsets of $\Omega_{2}$, a non empty subset $I$ of $\mathbb{R}$, a probability $P$ on $\Sigma$, a filtration $M$ of $I$ and $\Sigma$, and a stochastic process $S$ of $I, \Sigma, \Sigma_{2}$ and $P$. Suppose $S$ is $M$-stochastic process w.r.t. filtration. Then $S$ is an adapted stochastic process of $I, \Sigma, \Sigma_{2}, P$ and $S$.

## 6. Example for a Stochastic Process

Let $k_{1}, k_{2}$ be elements of $\mathbb{R}, \Omega$ be a non empty set, and $k$ be an element of $\Omega$. The functors: $\operatorname{Set} 1-\mathcal{R} \mathcal{V}\left(k_{1}, k_{2}, k\right)$ and $\operatorname{Set} 4-\mathcal{R} \mathcal{V}\left(k_{1}, k_{2}, k\right)$ yielding elements of $\mathbb{R}$ are defined by terms
(Def. 19) $\begin{cases}k_{1}, & \text { if } k=1 \text { or } k=2, \\ k_{2}, & \text { otherwise },\end{cases}$
(Def. 20) $\begin{cases}k_{1}, & \text { if } k=3, \\ k_{2}, & \text { otherwise },\end{cases}$
respectively. Let $k_{2}, k_{3}, k_{4}$ be elements of $\mathbb{R}$. The functor $\operatorname{Set} 3-\mathcal{R} \mathcal{V}\left(k_{2}, k_{3}, k_{4}, k\right)$ yielding an element of $\mathbb{R}$ is defined by the term
(Def. 21) $\begin{cases}k_{2}, & \text { if } k=2, \\ \operatorname{Set} 4-\mathcal{R} \mathcal{V}\left(k_{3}, k_{4}, k\right), & \text { otherwise. }\end{cases}$
Let $k_{1}, k_{2}, k_{3}, k_{4}$ be elements of $\mathbb{R}$. The functor $\operatorname{Set} 2-\mathcal{R} \mathcal{V}\left(k_{1}, k_{2}, k_{3}, k_{4}, k\right)$ yielding an element of $\mathbb{R}$ is defined by the term
(Def. 22) $\begin{cases}k_{1}, & \text { if } k=1, \\ \operatorname{Set} 3-\mathcal{R} \mathcal{V}\left(k_{2}, k_{3}, k_{4}, k\right), & \text { otherwise. }\end{cases}$
Now we state the proposition:
(35) Let us consider elements $k_{1}, k_{2}, k_{3}, k_{4}$ of $\mathbb{R}$, and a set $\Omega$. Suppose $\Omega=$ $\{1,2,3,4\}$. Then there exists a function $f$ from $\Omega$ into $\mathbb{R}$ such that
(i) $f(1)=k_{1}$, and
(ii) $f(2)=k_{2}$, and
(iii) $f(3)=k_{3}$, and
(iv) $f(4)=k_{4}$.

Proof: Define $\mathcal{U}$ (element of $\Omega$ ) $=\operatorname{Set} 2-\mathcal{R} \mathcal{V}\left(k_{1}, k_{2}, k_{3}, k_{4}, \$_{1}\right)$. Consider $f$ being a function from $\Omega$ into $\mathbb{R}$ such that for every element $d$ of $\Omega$, $f(d)=\mathcal{U}(d)$ from [5, Sch. 4]. $f(1)=k_{1} . f(2)=k_{2} . f(3)=k_{3} . f(4)=k_{4}$.

Let us consider a set $\Omega$.
Let us assume that $\Omega=\{1,2,3,4\}$. Now we state the propositions:
(36) There exists a function $f$ from $\Omega$ into $\mathbb{R}$ such that
(i) $f(1)=100$, and
(ii) $f(2)=100$, and
(iii) $f(3)=100$, and
(iv) $f(4)=100$.

The theorem is a consequence of (35).
(37) There exists a function $f$ from $\Omega$ into $\mathbb{R}$ such that
(i) $f(1)=80$, and
(ii) $f(2)=80$, and
(iii) $f(3)=120$, and
(iv) $f(4)=120$.

The theorem is a consequence of (35).
(38) There exists a function $f$ from $\Omega$ into $\mathbb{R}$ such that
(i) $f(1)=60$, and
(ii) $f(2)=80$, and
(iii) $f(3)=100$, and
(iv) $f(4)=120$.

The theorem is a consequence of (35).
(39) Let us consider elements $k_{1}, k_{2}, k_{3}, k_{4}$ of $\mathbb{R}$, and a non empty set $\Omega$. Suppose $\Omega=\{1,2,3,4\}$. Let us consider a $\sigma$-field $\Sigma$ of subsets of $\Omega$, a non empty subset $I$ of $\mathbb{R}$, and a filtration $M$ of $I$ and $\Sigma$. Suppose $M(1)=\Omega_{\text {now }}$ and $M(2)=\Omega_{f u t 1}$ and $M(3)=\Omega_{f u t 2}$. Let us consider an element $k$ of $I$. Suppose $k=3$. Then there exists a function $f$ from $\Omega$ into $\mathbb{R}$ such that
(i) $f(1)=k_{1}$, and
(ii) $f(2)=k_{2}$, and
(iii) $f(3)=k_{3}$, and
(iv) $f(4)=k_{4}$, and
(v) $f$ is random variable on the $k-\mathcal{E} \mathcal{F}$ of $M$ and the Borel sets.

Proof: Consider $f$ being a function from $\Omega$ into $\mathbb{R}$ such that $f(1)=k_{1}$ and $f(2)=k_{2}$ and $f(3)=k_{3}$ and $f(4)=k_{4} .1,2,3,4 \in \operatorname{dom} f . f$ is random variable on the $k-\mathcal{E} \mathcal{F}$ of $M$ and the Borel sets by [4, (1)], [11, (4)].

Let us consider a non empty set $\Omega$, a $\sigma$-field $\Sigma$ of subsets of $\Omega$, a non empty subset $I$ of $\mathbb{R}$, a filtration $M$ of $I$ and $\Sigma$, and an element $k$ of $I$.

Let us assume that $\Omega=\{1,2,3,4\}$. Now we state the propositions:
(40) Suppose $M(1)=\Omega_{\text {now }}$ and $M(2)=\Omega_{f u t 1}$ and $M(3)=\Omega_{f u t 2}$. Then suppose $k=3$. Then there exists a function $f$ from $\Omega$ into $\mathbb{R}$ such that
(i) $f(1)=60$, and
(ii) $f(2)=80$, and
(iii) $f(3)=100$, and
(iv) $f(4)=120$, and
(v) $f$ is random variable on the $k-\mathcal{E} \mathcal{F}$ of $M$ and the Borel sets.

The theorem is a consequence of (39).
(41) Suppose $M(1)=\Omega_{\text {now }}$ and $M(2)=\Omega_{f u t 1}$ and $M(3)=\Omega_{f u t 2}$. Then suppose $k=3$. Then there exists a function $f$ from $\Omega$ into $\mathbb{R}$ such that
(i) $f(1)=180$, and
(ii) $f(2)=120$, and
(iii) $f(3)=120$, and
(iv) $f(4)=80$, and
(v) $f$ is random variable on the $k-\mathcal{E} \mathcal{F}$ of $M$ and the Borel sets.

The theorem is a consequence of (39).
(42) Let us consider elements $k_{1}, k_{2}$ of $\mathbb{R}$, and a non empty set $\Omega$. Suppose $\Omega=\{1,2,3,4\}$. Let us consider a $\sigma$-field $\Sigma$ of subsets of $\Omega$, a non empty subset $I$ of $\mathbb{R}$, and a filtration $M$ of $I$ and $\Sigma$. Suppose $M(1)=\Omega_{\text {now }}$ and $M(2)=\Omega_{f u t 1}$ and $M(3)=\Omega_{f u t 2}$. Let us consider an element $k$ of $I$. Suppose $k=2$. Then there exists a function $f$ from $\Omega$ into $\mathbb{R}$ such that
(i) $f(1)=k_{1}$, and
(ii) $f(2)=k_{1}$, and
(iii) $f(3)=k_{2}$, and
(iv) $f(4)=k_{2}$, and
(v) $f$ is random variable on the $k-\mathcal{E} \mathcal{F}$ of $M$ and the Borel sets.

Proof: Consider $f$ being a function from $\Omega$ into $\mathbb{R}$ such that $f(1)=k_{1}$ and $f(2)=k_{1}$ and $f(3)=k_{2}$ and $f(4)=k_{2}$. Set $i=k$. For every set $x$, $f^{-1}(x) \in$ the $i-\mathcal{E} \mathcal{F}$ of $M$ by [4, (1)].
Let us consider a non empty set $\Omega$, a $\sigma$-field $\Sigma$ of subsets of $\Omega$, a non empty subset $I$ of $\mathbb{R}$, a filtration $M$ of $I$ and $\Sigma$, and an element $k$ of $I$.

Let us assume that $\Omega=\{1,2,3,4\}$. Now we state the propositions:
(43) Suppose $M(1)=\Omega_{\text {now }}$ and $M(2)=\Omega_{f u t 1}$ and $M(3)=\Omega_{f u t 2}$. Then suppose $k=2$. Then there exists a function $f$ from $\Omega$ into $\mathbb{R}$ such that
(i) $f(1)=80$, and
(ii) $f(2)=80$, and
(iii) $f(3)=120$, and
(iv) $f(4)=120$, and
(v) $f$ is random variable on the $k-\mathcal{E} \mathcal{F}$ of $M$ and the Borel sets.

The theorem is a consequence of (42).
(44) Suppose $M(1)=\Omega_{\text {now }}$ and $M(2)=\Omega_{f u t 1}$ and $M(3)=\Omega_{f u t 2}$. Then suppose $k=2$. Then there exists a function $f$ from $\Omega$ into $\mathbb{R}$ such that
(i) $f(1)=150$, and
(ii) $f(2)=150$, and
(iii) $f(3)=100$, and
(iv) $f(4)=100$, and
(v) $f$ is random variable on the $k-\mathcal{E} \mathcal{F}$ of $M$ and the Borel sets.

The theorem is a consequence of (42).
(45) Let us consider an element $k_{1}$ of $\mathbb{R}$, and a non empty set $\Omega$. Suppose $\Omega=\{1,2,3,4\}$. Let us consider a $\sigma$-field $\Sigma$ of subsets of $\Omega$, a non empty subset $I$ of $\mathbb{R}$, and a filtration $M$ of $I$ and $\Sigma$. Suppose $M(1)=\Omega_{\text {now }}$ and $M(2)=\Omega_{f u t 1}$ and $M(3)=\Omega_{f u t 2}$. Let us consider an element $k$ of $I$. Suppose $k=1$. Then there exists a function $f$ from $\Omega$ into $\mathbb{R}$ such that
(i) $f(1)=k_{1}$, and
(ii) $f(2)=k_{1}$, and
(iii) $f(3)=k_{1}$, and
(iv) $f(4)=k_{1}$, and
(v) $f$ is random variable on the $k-\mathcal{E} \mathcal{F}$ of $M$ and the Borel sets.

Proof: Consider $f$ being a function from $\Omega$ into $\mathbb{R}$ such that $f(1)=k_{1}$ and $f(2)=k_{1}$ and $f(3)=k_{1}$ and $f(4)=k_{1}$. Set $i=k$. For every set $x$ such that $x \in$ the Borel sets holds $f^{-1}(x) \in$ the $i-\mathcal{E \mathcal { F }}$ of $M$ by [4, (1)].

Let us consider a non empty set $\Omega$, a $\sigma$-field $\Sigma$ of subsets of $\Omega$, a non empty subset $I$ of $\mathbb{R}$, a filtration $M$ of $I$ and $\Sigma$, and an element $k$ of $I$.

Let us assume that $\Omega=\{1,2,3,4\}$. Now we state the propositions:
(46) Suppose $M(1)=\Omega_{\text {now }}$ and $M(2)=\Omega_{f u t 1}$ and $M(3)=\Omega_{f u t 2}$. Then suppose $k=1$. Then there exists a function $f$ from $\Omega$ into $\mathbb{R}$ such that
(i) $f(1)=100$, and
(ii) $f(2)=100$, and
(iii) $f(3)=100$, and
(iv) $f(4)=100$, and
(v) $f$ is random variable on the $k-\mathcal{E} \mathcal{F}$ of $M$ and the Borel sets.

The theorem is a consequence of (45).
(47) Suppose $M(1)=\Omega_{\text {now }}$ and $M(2)=\Omega_{f u t 1}$ and $M(3)=\Omega_{f u t 2}$. Then suppose $k=1$. Then there exists a function $f$ from $\Omega$ into $\mathbb{R}$ such that
(i) $f(1)=125$, and
(ii) $f(2)=125$, and
(iii) $f(3)=125$, and
(iv) $f(4)=125$, and
(v) $f$ is random variable on the $k-\mathcal{E} \mathcal{F}$ of $M$ and the Borel sets.

The theorem is a consequence of (45).
Now we state the proposition:
(48) Let us consider a non empty set $\Omega$. Suppose $\Omega=\{1,2,3,4\}$. Let us consider a $\sigma$-field $\Sigma$ of subsets of $\Omega$, and a non empty subset $I$ of $\mathbb{R}$. Suppose $I=\{1,2,3\}$ and $\Sigma=2^{\{1,2,3,4\}}$. Let us consider a filtration $M$ of $I$ and $\Sigma$. Suppose $M(1)=\Omega_{\text {now }}$ and $M(2)=\Omega_{f u t 1}$ and $M(3)=\Omega_{f u t 2}$. Let us consider a probability $P$ on $\Sigma$, and an element $i$ of $I$. Then there exists a function $\mathcal{R} \mathcal{V}$ from $\Omega$ into $\mathbb{R}$ such that $\mathcal{R} \mathcal{V}$ is random variable on the $i-\mathcal{E} \mathcal{F}$ of $M$ and the Borel sets. The theorem is a consequence of (46), (43), and (40).

Let $I$ be a non empty subset of $\mathbb{R}$. Assume $I=\{1,2,3\}$. Let $i$ be an element of $I$. Assume $i=2$ or $i=3$. Let $\Omega$ be a non empty set. Assume $\Omega=\{1,2,3,4\}$. Let $\Sigma$ be a $\sigma$-field of subsets of $\Omega$. Assume $\Sigma=2^{\Omega}$. Let $f_{1}$ be a function from $\Omega$ into $\mathbb{R}$. Assume $f_{1}(1)=60$ and $f_{1}(2)=80$ and $f_{1}(3)=100$ and $f_{1}(4)=120$. Let $f_{2}$ be a function from $\Omega$ into $\mathbb{R}$. Assume $f_{2}(1)=80$ and $f_{2}(2)=80$ and $f_{2}(3)=120$ and $f_{2}(4)=120$. Let $f_{3}$ be a function from $\Omega$ into $\mathbb{R}$. The functor Select12- $\mathcal{R} \mathcal{V}\left(i, \Sigma, f_{1}, f_{2}, f_{3}\right)$ yielding an element of the set of random variables on $\Sigma$ and the Borel sets is defined by the term
(Def. 23)

$$
\begin{cases}f_{2}, & \text { if } i=2 \\ f_{1}, & \text { otherwise }\end{cases}
$$

Assume $I=\{1,2,3\}$. Assume $\Omega=\{1,2,3,4\}$. Assume $\Sigma=2^{\Omega}$. Let $f_{1}, f_{2}$ be functions from $\Omega$ into $\mathbb{R}$. Assume $f_{3}(1)=100$ and $f_{3}(2)=100$ and $f_{3}(3)=100$ and $f_{3}(4)=100$. The functor Select123- $\mathcal{R} \mathcal{V}\left(i, \Sigma, f_{1}, f_{2}, f_{3}\right)$ yielding an element of the set of random variables on $\Sigma$ and the Borel sets is defined by the term
(Def. 24)

$$
\begin{cases}\text { Select12- } \mathcal{R} \mathcal{V}\left(i, \Sigma, f_{1}, f_{2}, f_{3}\right), & \text { if } i=2 \text { or } i=3 \\ f_{3}, & \text { otherwise }\end{cases}
$$

Now we state the proposition:
(49) Let us consider non empty sets $\Omega, \Omega_{2}$. Suppose $\Omega=\{1,2,3,4\}$. Let us consider a $\sigma$-field $\Sigma$ of subsets of $\Omega$, and a non empty subset $I$ of $\mathbb{R}$. Suppose $I=\{1,2,3\}$ and $\Sigma=2^{\{1,2,3,4\}}$. Let us consider a probability $P$ on $\Sigma$, and a filtration $M$ of $I$ and $\Sigma$. Suppose $M(1)=\Omega_{\text {now }}$ and $M(2)=\Omega_{\text {fut } 1}$ and $M(3)=\Omega_{f u t 2}$. Then there exists a stochastic process $S$ of $I, \Sigma$, the Borel sets and $P$ such that
(i) for every element $k$ of $I$, there exists a function $\mathcal{R} \mathcal{V}$ from $\Omega$ into $\mathbb{R}$ such that $S(k)=\mathcal{R} \mathcal{V}$ and $\mathcal{R} \mathcal{V}$ is random variable on $\Sigma$ and the Borel sets and random variable on the $k-\mathcal{E} \mathcal{F}$ of $M$ and the Borel sets and there exists a function $f$ from $\Omega$ into $\mathbb{R}$ such that if $k=1$, then $f(1)=100$ and $f(2)=100$ and $f(3)=100$ and $f(4)=100$ and $S(k)=f$ and there exists a function $f$ from $\Omega$ into $\mathbb{R}$ such that if $k=2$, then $f(1)=80$ and $f(2)=80$ and $f(3)=120$ and $f(4)=120$ and $S(k)=f$ and there exists a function $f$ from $\Omega$ into $\mathbb{R}$ such that if $k=3$, then $f(1)=60$ and $f(2)=80$ and $f(3)=100$ and $f(4)=120$ and $S(k)=f$ and $S$ is $M$-stochastic process w.r.t. filtration, and
(ii) $S$ is an adapted stochastic process of $I, \Sigma$, the Borel sets, $P$ and $S$.

Proof: Consider $f_{3}$ being a function from $\Omega$ into $\mathbb{R}$ such that $f_{3}(1)=$ 100 and $f_{3}(2)=100$ and $f_{3}(3)=100$ and $f_{3}(4)=100$. Consider $f_{2}$ being a function from $\Omega$ into $\mathbb{R}$ such that $f_{2}(1)=80$ and $f_{2}(2)=80$ and $f_{2}(3)=120$ and $f_{2}(4)=120$. Consider $f_{1}$ being a function from $\Omega$ into $\mathbb{R}$ such that $f_{1}(1)=60$ and $f_{1}(2)=80$ and $f_{1}(3)=100$ and $f_{1}(4)=120$.
 a function from $I$ into the set of random variables on $\Sigma$ and the Borel sets such that for every element $d$ of $I, f_{4}(d)=\mathcal{U}(d)$ from [5, Sch. 4]. For every element $k$ of $I$, there exists a function $\mathcal{R} \mathcal{V}$ from $\Omega$ into $\mathbb{R}$ such that $f_{4}(k)=\mathcal{R} \mathcal{V}$ and $\mathcal{R} \mathcal{V}$ is random variable on $\Sigma$ and the Borel sets. For every element $k$ of $I$, there exists a function $\mathcal{R} \mathcal{V}$ from $\Omega$ into $\mathbb{R}$ such that $f_{4}(k)=\mathcal{R} \mathcal{V}$ and $\mathcal{R} \mathcal{V}$ is random variable on $\Sigma$ and the Borel sets and random variable on the $k-\mathcal{E} \mathcal{F}$ of $M$ and the Borel sets and there exists
a function $f$ from $\Omega$ into $\mathbb{R}$ such that if $k=1$, then $f(1)=100$ and $f(2)=100$ and $f(3)=100$ and $f(4)=100$ and $f_{4}(k)=f$ and there exists a function $f$ from $\Omega$ into $\mathbb{R}$ such that if $k=2$, then $f(1)=80$ and $f(2)=80$ and $f(3)=120$ and $f(4)=120$ and $f_{4}(k)=f$ and there exists a function $f$ from $\Omega$ into $\mathbb{R}$ such that if $k=3$, then $f(1)=60$ and $f(2)=80$ and $f(3)=100$ and $f(4)=120$ and $f_{4}(k)=f$ and $f_{4}$ is $M$-stochastic process w.r.t. filtration and adapted stochastic process of $I$, $\Sigma$, the Borel sets, $P$ and $f_{4}$.

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# Circumcenter，Circumcircle and Centroid of a Triangle 

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#### Abstract

Summary．We introduce，using the Mizar system［1］，some basic concepts of Euclidean geometry：the half length and the midpoint of a segment，the per－ pendicular bisector of a segment，the medians（the cevians that join the vertices of a triangle to the midpoints of the opposite sides）of a triangle．

We prove the existence and uniqueness of the circumcenter of a triangle（the intersection of the three perpendicular bisectors of the sides of the triangle）．The extended law of sines and the formula of the radius of the Morley＇s trisector triangle are formalized［3］．

Using the generalized Ceva＇s Theorem，we prove the existence and uniqueness of the centroid（the common point of the medians［4）of a triangle．


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## 1．Preliminaries

From now on $n$ denotes a natural number，$\lambda, \lambda_{2}, \mu, \mu_{2}$ denote real numbers， $x_{1}, x_{2}$ denote elements of $\mathcal{R}^{n}, A_{1}, B_{1}, C_{1}$ denote points of $\mathcal{E}_{\mathrm{T}}^{n}$ ，and $a$ denotes a real number．

Now we state the propositions：
（1）If $A_{1}=(1-\lambda) \cdot x_{1}+\lambda \cdot x_{2}$ and $B_{1}=(1-\mu) \cdot x_{1}+\mu \cdot x_{2}$ ，then $B_{1}-A_{1}=$ $(\mu-\lambda) \cdot\left(x_{2}-x_{1}\right)$ ．
（2）If $|a|=|1-a|$ ，then $a=\frac{1}{2}$ ．

In the sequel $P, A, B$ denote elements of $\mathcal{R}^{n}$ and $L$ denotes an element of Lines $\left(\mathcal{R}^{n}\right)$.

Now we state the propositions:
(3) Line $(P, P)=\{P\}$.
(4) If $A_{1}=A$ and $B_{1}=B$, then $\operatorname{Line}\left(A_{1}, B_{1}\right)=\operatorname{Line}(A, B)$.
(5) If $A_{1} \neq C_{1}$ and $C_{1} \in \mathcal{L}\left(A_{1}, B_{1}\right)$ and $A_{1}, C_{1} \in L$ and $L$ is a line, then $B_{1} \in L$. The theorem is a consequence of (4).
Let $n$ be a natural number and $S$ be a subset of $\mathcal{R}^{n}$. We say that $S$ is a point if and only if
(Def. 1) there exists an element $P$ of $\mathcal{R}^{n}$ such that $S=\{P\}$.
Now we state the propositions:
(6) (i) $L$ is a line, or
(ii) there exists an element $P$ of $\mathcal{R}^{n}$ such that $L=\{P\}$.

The theorem is a consequence of (3).
(7) $L$ is a line or a point.

Let us assume that $L$ is a line. Now we state the propositions:
(8) There exists no element $P$ of $\mathcal{R}^{n}$ such that $L=\{P\}$.
(9) $L$ is not a point.

## 2. Betweenness

In the sequel $A, B, C$ denote points of $\mathcal{E}_{\mathrm{T}}^{2}$.
Now we state the propositions:
(10) If $C \in \mathcal{L}(A, B)$, then $|A-B|=|A-C|+|C-B|$.
(11) If $|A-B|=|A-C|+|C-B|$, then $C \in \mathcal{L}(A, B)$. The theorem is a consequence of (10).
(12) Let us consider points $p, q_{1}, q_{2}$ of $\mathcal{E}_{\mathrm{T}}^{2}$. Then $p \in \mathcal{L}\left(q_{1}, q_{2}\right)$ if and only if $\rho\left(q_{1}, p\right)+\rho\left(p, q_{2}\right)=\rho\left(q_{1}, q_{2}\right)$. The theorem is a consequence of (11).
Let us consider elements $p, q, r$ of $\mathcal{E}^{2}$.
Let us assume that $p, q, r$ are mutually different and $p=A$ and $q=B$ and $r=C$. Now we state the propositions:
(13) $\quad A \in \mathcal{L}(B, C)$ if and only if $p$ is between $q$ and $r$. The theorem is a consequence of (12) and (11).
(14) $A \in \mathcal{L}(B, C)$ if and only if $p$ is between $q$ and $r$. The theorem is a consequence of (13).

## 3. Real Plane

From now on $x, y, z, y_{1}, y_{2}$ denote elements of $\mathcal{R}^{2}, L, L_{1}, L_{2}$ denote elements of Lines $\left(\mathcal{R}^{2}\right), D, E, F$ denote points of $\mathcal{E}_{\mathrm{T}}^{2}$, and $b, c, d, r, s$ denote real numbers.

Now we state the propositions:
(15) Let us consider elements $O, O_{1}, O_{2}$ of $\mathcal{R}^{2}$. Suppose $O=[0,0]$ and $O_{1}=$ $[1,0]$ and $O_{2}=[0,1]$. Then $\mathcal{R}^{2}=\operatorname{Plane}\left(O, O_{1}, O_{2}\right)$.
(16) $\mathcal{R}^{2}$ is an element of $\operatorname{Planes}\left(\mathcal{R}^{2}\right)$. The theorem is a consequence of (15).
(17) (i) $[1,0] \neq[0,1]$, and
(ii) $[1,0] \neq[0,0]$, and
(iii) $[0,1] \neq[0,0]$.
(18) There exists $x$ such that $x \notin L$. The theorem is a consequence of (6) and (17).
(19) There exists $L$ such that
(i) $L$ is a point, and
(ii) $L$ misses $L_{1}$.

The theorem is a consequence of (18) and (3).
Let us assume that $L_{1} \nVdash L_{2}$. Now we state the propositions:
(20) (i) there exists $x$ such that $L_{1}=\{x\}$ or $L_{2}=\{x\}$, or
(ii) $L_{1}$ is a line and $L_{2}$ is a line and there exists $x$ such that $L_{1} \cap L_{2}=\{x\}$.

The theorem is a consequence of (3) and (16).
(21) (i) $L_{1}$ is a point, or
(ii) $L_{2}$ is a point, or
(iii) $L_{1}$ is a line and $L_{2}$ is a line and $L_{1} \cap L_{2}$ is a point.

Now we state the proposition:
(22) If $L_{1} \cap L_{2}$ is a point and $A \in L_{1} \cap L_{2}$, then $L_{1} \cap L_{2}=\{A\}$.

## 4. The Midpoint of a Segment

Let $A, B$ be points of $\mathcal{E}_{\mathrm{T}}^{2}$. The functor half-length $(A, B)$ yielding a real number is defined by the term
(Def. 2) $\quad\left(\frac{1}{2}\right) \cdot|A-B|$.
Now we state the propositions:
(23) half-length $(A, B)=$ half-length $(B, A)$.
(24) half-length $(A, A)=0$.
(25) $\left|A-\left(\frac{1}{2}\right) \cdot(A+B)\right|=\left(\frac{1}{2}\right) \cdot|A-B|$.
(26) There exists $C$ such that
(i) $C \in \mathcal{L}(A, B)$, and
(ii) $|A-C|=\left(\frac{1}{2}\right) \cdot|A-B|$.

The theorem is a consequence of (25).
(27) If $|A-B|=|A-C|$ and $B, C \in \mathcal{L}(A, D)$, then $B=C$. The theorem is a consequence of (1).
Let $A, B$ be points of $\mathcal{E}_{\mathrm{T}}^{2}$. The functor $\operatorname{SegMidpoint}(A, B)$ yielding a point of $\mathcal{E}_{\mathrm{T}}^{2}$ is defined by
(Def. 3) there exists $C$ such that $C \in \mathcal{L}(A, B)$ and it $=C$ and $|A-C|=$ half-length $(A, B)$.
Now we state the propositions:
(28) $\operatorname{SegMidpoint}(A, B) \in \mathcal{L}(A, B)$.
(29) SegMidpoint $(A, B)=\left(\frac{1}{2}\right) \cdot(A+B)$. The theorem is a consequence of (25).
(30) $\operatorname{SegMidpoint}(A, B)=\operatorname{SegMidpoint}(B, A)$. The theorem is a consequence of (29).
(31) $\operatorname{SegMidpoint}(A, A)=A$. The theorem is a consequence of (29).
(32) If $\operatorname{SegMidpoint}(A, B)=A$, then $A=B$. The theorem is a consequence of (29).
(33) If $\operatorname{SegMidpoint}(A, B)=B$, then $A=B$. The theorem is a consequence of (30) and (32).
Let us assume that $C \in \mathcal{L}(A, B)$ and $|A-C|=|B-C|$. Now we state the propositions:
(34) half-length $(A, B)=|A-C|$. The theorem is a consequence of (10).
(35) $C=\operatorname{SegMidpoint}(A, B)$. The theorem is a consequence of (34).

Now we state the propositions:
(36) $\quad|A-\operatorname{SegMidpoint}(A, B)|=|\operatorname{SegMidpoint}(A, B)-B|$. The theorem is a consequence of (29) and (25).
(37) If $A \neq B$ and $r$ is positive and $r \neq 1$ and $|A-C|=r \cdot|A-B|$, then $A$, $B, C$ are mutually different.
(38) If $C \in \mathcal{L}(A, B)$ and $|A-C|=\left(\frac{1}{2}\right) \cdot|A-B|$, then $|B-C|=\left(\frac{1}{2}\right) \cdot|A-B|$. The theorem is a consequence of (10).

## 5. Perpendicularity

Now we state the propositions:
(39) $\quad L_{1}$ and $L_{2}$ are coplanar. The theorem is a consequence of (15).
(40) If $L_{1} \perp L_{2}$, then $L_{1}$ meets $L_{2}$.
(41) If $L_{1}$ is a line and $L_{2}$ is a line and $L_{1}$ misses $L_{2}$, then $L_{1} \| L_{2}$.
(42) Suppose $L_{1} \neq L_{2}$ and $L_{1}$ meets $L_{2}$. Then
(i) there exists $x$ such that $L_{1}=\{x\}$ or $L_{2}=\{x\}$, or
(ii) $L_{1}$ is a line and $L_{2}$ is a line and there exists $x$ such that $L_{1} \cap L_{2}=\{x\}$.

The theorem is a consequence of (20).
Let us assume that $L_{1} \perp L_{2}$. Now we state the propositions:
(43) There exists $x$ such that $L_{1} \cap L_{2}=\{x\}$. The theorem is a consequence of (39), (8), and (42).
(44) $L_{1} \cap L_{2}$ is a point.

Now we state the propositions:
(45) If $L_{1} \perp L_{2}$, then $L_{1} \nVdash L_{2}$. The theorem is a consequence of (39).
(46) If $L_{1}$ is a line and $L_{2}$ is a line and $L_{1} \| L_{2}$, then $L_{1} \not \perp L_{2}$.

Now we state the propositions:
(47) If $L_{1}$ is a line, then there exists $L_{2}$ such that $x \in L_{2}$ and $L_{1} \perp L_{2}$. The theorem is a consequence of (18).
(48) If $L_{1} \perp L_{2}$ and $L_{1}=\operatorname{Line}(A, B)$ and $L_{2}=\operatorname{Line}(C, D)$, then $\mid(B-A, D-$ $C) \mid=0$. The theorem is a consequence of (1).
(49) If $L$ is a line and $A, B \in L$ and $A \neq B$, then $L=\operatorname{Line}(A, B)$. The theorem is a consequence of (4).
Let us assume that $L_{1} \perp L_{2}$ and $C \in L_{1} \cap L_{2}$ and $A \in L_{1}$ and $B \in L_{2}$ and $A \neq C$ and $B \neq C$. Now we state the propositions:
(50) (i) $\measuredangle(A, C, B)=\frac{\pi}{2}$, or
(ii) $\measuredangle(A, C, B)=\frac{3 \cdot \pi}{2}$.

The theorem is a consequence of (49) and (48).
(51) $A, B, C$ form a triangle.

Proof: $A \notin \operatorname{Line}(B, C)$ by [5, (67)], (43), (49).

## 6. The Perpendicular Bisector of a Segment

Now we state the proposition:
(52) Suppose $A \neq B$ and $L_{1}=\operatorname{Line}(A, B)$ and $C \in \mathcal{L}(A, B)$ and $|A-C|=$ $\left(\frac{1}{2}\right) \cdot|A-B|$. Then there exists $L_{2}$ such that
(i) $C \in L_{2}$, and
(ii) $L_{1} \perp L_{2}$.

The theorem is a consequence of (4) and (47).
Let $A, B$ be elements of $\mathcal{E}_{\mathrm{T}}^{2}$. Assume $A \neq B$. The functor $\operatorname{PerpBisec}(A, B)$ yielding an element of $\operatorname{Lines}\left(\mathcal{R}^{2}\right)$ is defined by
(Def. 4) there exist elements $L_{1}, L_{2}$ of $\operatorname{Lines}\left(\mathcal{R}^{2}\right)$ such that it $=L_{2}$ and $L_{1}=$ Line $(A, B)$ and $L_{1} \perp L_{2}$ and $L_{1} \cap L_{2}=\{\operatorname{SegMidpoint}(A, B)\}$.
Let us assume that $A \neq B$. Now we state the propositions:
(53) $\operatorname{PerpBisec}(A, B)$ is a line.
(54) $\operatorname{PerpBisec}(A, B)=\operatorname{PerpBisec}(B, A)$. The theorem is a consequence of (43), (16), and (30).
(55) Suppose $A \neq B$ and $L_{1}=\operatorname{Line}(A, B)$ and $C \in \mathcal{L}(A, B)$ and $|A-C|=$ $\left(\frac{1}{2}\right) \cdot|A-B|$ and $C \in L_{2}$ and $L_{1} \perp L_{2}$ and $D \in L_{2}$. Then $|D-A|=|D-B|$. The theorem is a consequence of (38), (37), and (50).
(56) If $A \neq B$ and $C \in \operatorname{PerpBisec}(A, B)$, then $|C-A|=|C-B|$. The theorem is a consequence of (28) and (55).
(57) If $C \in \operatorname{Line}(A, B)$ and $|A-C|=|B-C|$, then $C \in \mathcal{L}(A, B)$. The theorem is a consequence of $(4),(3)$, and (2).
(58) If $A \neq B$, then $\operatorname{SegMidpoint}(A, B) \in \operatorname{PerpBisec}(A, B)$.
(59) If $A \neq B$ and $L_{1}=\operatorname{Line}(A, B)$ and $L_{1} \perp L_{2}$ and $\operatorname{SegMidpoint}(A, B) \in$ $L_{2}$, then $L_{2}=\operatorname{PerpBisec}(A, B)$. The theorem is a consequence of (16).
(60) If $A \neq B$ and $|C-A|=|C-B|$, then $C \in \operatorname{PerpBisec}(A, B)$. The theorem is a consequence of $(47),(43),(50),(57),(35),(58)$, and (59).

## 7. The Circumcircle of a Triangle

Let us assume that $A, B, C$ form a triangle. Now we state the propositions:
(61) $\operatorname{PerpBisec}(A, B) \cap \operatorname{PerpBisec}(B, C)$ is a point. The theorem is a consequence of (16), (8), and (20).
(62) There exists $D$ such that
(i) $\operatorname{PerpBisec}(A, B) \cap \operatorname{PerpBisec}(B, C)=\{D\}$, and
(ii) $\operatorname{PerpBisec}(B, C) \cap \operatorname{PerpBisec}(C, A)=\{D\}$, and
(iii) $\operatorname{PerpBisec}(C, A) \cap \operatorname{PerpBisec}(A, B)=\{D\}$, and
(iv) $|D-A|=|D-B|$, and
(v) $|D-A|=|D-C|$, and
(vi) $|D-B|=|D-C|$.

The theorem is a consequence of (61), (56), and (60).
Let $A, B, C$ be points of $\mathcal{E}_{\mathrm{T}}^{2}$. Assume $A, B, C$ form a triangle. The functor Circumcenter $\triangle(A, B, C)$ yielding a point of $\mathcal{E}_{\mathrm{T}}^{2}$ is defined by
(Def. 5) $\operatorname{PerpBisec}(A, B) \cap \operatorname{PerpBisec}(B, C)=\{i t\}$ and
$\operatorname{PerpBisec}(B, C) \cap \operatorname{PerpBisec}(C, A)=\{i t\}$ and
$\operatorname{PerpBisec}(C, A) \cap \operatorname{PerpBisec}(A, B)=\{i t\}$.
Assume $A, B, C$ form a triangle. The functor RadCircumCirc $\triangle(A, B, C)$ yielding a real number is defined by the term
(Def. 6) $\mid$ Circumcenter $\triangle(A, B, C)-A \mid$.
(63) If $A, B, C$ form a triangle, then there exists $a$ and there exists $b$ and there exists $r$ such that $A, B, C \in \operatorname{circle}(a, b, r)$. The theorem is a consequence of (62).
(64) Suppose $A, B, C$ form a triangle and $A, B, C \in \operatorname{circle}(a, b, r)$. Then
(i) $[a, b]=$ Circumcenter $\triangle(A, B, C)$, and
(ii) $r=\mid$ Circumcenter $\triangle(A, B, C)-A \mid$.

The theorem is a consequence of (60), (22), and (61).
Let us assume that $A, B, C$ form a triangle. Now we state the propositions:
(65) RadCircumCirc $\triangle(A, B, C)>0$. The theorem is a consequence of (63) and (64).
(66) (i) $\mid$ Circumcenter $\triangle(A, B, C)-A|=|$ Circumcenter $\triangle(A, B, C)-B \mid$, and
(ii) $\mid$ Circumcenter $\triangle(A, B, C)-A|=|$ Circumcenter $\triangle(A, B, C)-C \mid$, and
(iii) $\mid$ Circumcenter $\triangle(A, B, C)-B|=|$ Circumcenter $\triangle(A, B, C)-C \mid$.

The theorem is a consequence of (62).
(67) (i) RadCircumCirc $\triangle(A, B, C)=\mid$ Circumcenter $\triangle(A, B, C)-B \mid$, and
(ii) RadCircumCirc $\triangle(A, B, C)=\mid$ Circumcenter $\triangle(A, B, C)-C \mid$.

The theorem is a consequence of (66).
(68) If $A, B, C$ form a triangle and $A, B, C \in \operatorname{circle}(a, b, r)$ and $A, B$, $C \in \operatorname{circle}(c, d, s)$, then $a=c$ and $b=d$ and $r=s$. The theorem is a consequence of (64).
(69) If $r \neq s$, then $\operatorname{circle}(a, b, r)$ misses $\operatorname{circle}(a, b, s)$.

## 8. Extended Law of Sines

Now we state the propositions:
(70) Suppose $A, B, C$ form a triangle and $A, B, C \in \operatorname{circle}(a, b, r)$ and $A, B$, $D$ form a triangle and $A, B, D \in \operatorname{circle}(a, b, r)$ and $C \neq D$. Then
(i) $\varnothing_{0}(A, B, C)=\varnothing_{\cap}(D, B, C)$, or
(ii) $\varnothing_{0}(A, B, C)=-\varnothing_{0}(D, B, C)$.

Proof: $D, B, C$ form a triangle by [6, (20), (11)], [2, (68)], [6, (30)].
(71) Suppose $A, B, C$ form a triangle and $A, B, C \in \operatorname{circle}(a, b, r)$. Then
(i) $\varnothing_{\bigcirc}(A, B, C)=2 \cdot r$, or
(ii) $\varnothing_{0}(A, B, C)=-2 \cdot r$.

The theorem is a consequence of (70).
(72) If $A, B, C$ form a triangle and $0<\measuredangle(C, B, A)<\pi$, then $\varnothing_{\Omega}(A, B, C)>$ 0.
(73) If $A, B, C$ form a triangle and $\pi<\measuredangle(C, B, A)<2 \cdot \pi$, then $\varnothing_{\rho}(A, B, C)<$ 0.
(74) Suppose $A, B, C$ form a triangle and $0<\measuredangle(C, B, A)<\pi$ and $A, B$, $C \in \operatorname{circle}(a, b, r)$. Then $\varnothing_{\bigcirc}(A, B, C)=2 \cdot r$. The theorem is a consequence of (71) and (72).
(75) Suppose $A, B, C$ form a triangle and $\pi<\measuredangle(C, B, A)<2 \cdot \pi$ and $A, B$, $C \in \operatorname{circle}(a, b, r)$. Then $\varnothing_{0}(A, B, C)=-2 \cdot r$. The theorem is a consequence of (71) and (73).
(76) Suppose $A, B, C$ form a triangle and $0<\measuredangle(C, B, A)<\pi$ and $A, B$, $C \in \operatorname{circle}(a, b, r)$. Then
(i) $|A-B|=2 \cdot r \cdot \sin \measuredangle(A, C, B)$, and
(ii) $|B-C|=2 \cdot r \cdot \sin \measuredangle(B, A, C)$, and
(iii) $|C-A|=2 \cdot r \cdot \sin \measuredangle(C, B, A)$.

The theorem is a consequence of (74).
(77) Suppose $A, B, C$ form a triangle and $\pi<\measuredangle(C, B, A)<2 \cdot \pi$ and $A, B$, $C \in \operatorname{circle}(a, b, r)$. Then
(i) $|A-B|=-2 \cdot r \cdot \sin \measuredangle(A, C, B)$, and
(ii) $|B-C|=-2 \cdot r \cdot \sin \measuredangle(B, A, C)$, and
(iii) $|C-A|=-2 \cdot r \cdot \sin \measuredangle(C, B, A)$.

The theorem is a consequence of (75).
(78) Extended Law of Sines:

Suppose $A, B, C$ form a triangle and $0<\measuredangle(C, B, A)<\pi$ and $A, B$, $C \in \operatorname{circle}(a, b, r)$. Then
(i) $\frac{|A-B|}{\sin \measuredangle(A, C, B)}=2 \cdot r$, and
(ii) $\frac{|B-C|}{\sin \measuredangle(B, A, C)}=2 \cdot r$, and
(iii) $\frac{|C-A|}{\sin \measuredangle(C, B, A)}=2 \cdot r$.

The theorem is a consequence of (76).
(79) Suppose $A, B, C$ form a triangle and $\pi<\measuredangle(C, B, A)<2 \cdot \pi$ and $A, B$, $C \in \operatorname{circle}(a, b, r)$. Then
(i) $\frac{|A-B|}{\sin \measuredangle(A, C, B)}=-2 \cdot r$, and
(ii) $\frac{|B-C|}{\sin \measuredangle(B, A, C)}=-2 \cdot r$, and
(iii) $\frac{|C-A|}{\sin \measuredangle(C, B, A)}=-2 \cdot r$.

The theorem is a consequence of (77).

## 9. The Centroid of a Triangle

Now we state the proposition:
(80) Suppose $A, B, C$ form a triangle and $D=\left(1-\left(\frac{1}{2}\right)\right) \cdot B+\left(\frac{1}{2}\right) \cdot C$ and $E=\left(1-\left(\frac{1}{2}\right)\right) \cdot C+\left(\frac{1}{2}\right) \cdot A$ and $F=\left(1-\left(\frac{1}{2}\right)\right) \cdot A+\left(\frac{1}{2}\right) \cdot B$. Then Line $(A, D)$, Line $(B, E)$, Line $(C, F)$ are concurrent.
Let $A, B, C$ be points of $\mathcal{E}_{\mathrm{T}}^{2}$. The functor Median $\triangle(A, B, C)$ yielding an element of $\operatorname{Lines}\left(\mathcal{R}^{2}\right)$ is defined by the term
(Def. 7) Line $(A, \operatorname{SegMidpoint}(B, C))$.
(81) Median $\triangle(A, A, A)=\{A\}$. The theorem is a consequence of (4), (3), and (31).
(82) Median $\triangle(A, A, B)=\operatorname{Line}(A, B)$. The theorem is a consequence of (28), (32), (4), (3), and (81).
(83) Median $\triangle(A, B, A)=\operatorname{Line}(A, B)$. The theorem is a consequence of (28), (33), (4), (3), and (81).
(84) Median $\triangle(B, A, A)=\operatorname{Line}(A, B)$.

Let us assume that $A, B, C$ form a triangle. Now we state the propositions:
(85) Median $\triangle(A, B, C)$ is a line. The theorem is a consequence of (6) and (28).
(86) There exists $D$ such that
(i) $D \in$ Median $\triangle(A, B, C)$, and
(ii) $D \in \operatorname{Median} \triangle(B, C, A)$, and
(iii) $D \in \operatorname{Median} \triangle(C, A, B)$.

The theorem is a consequence of (29), (80), and (4).
(87) There exists $D$ such that
(i) Median $\triangle(A, B, C) \cap$ Median $\triangle(B, C, A)=\{D\}$, and
(ii) Median $\triangle(B, C, A) \cap$ Median $\triangle(C, A, B)=\{D\}$, and
(iii) Median $\triangle(C, A, B) \cap$ Median $\triangle(A, B, C)=\{D\}$.

The theorem is a consequence of (86), (4), (85), (28), (32), (5), (8), and (20).

Let $A, B, C$ be points of $\mathcal{E}_{\mathrm{T}}^{2}$. Assume $A, B, C$ form a triangle. The functor Centroid $\triangle(A, B, C)$ yielding a point of $\mathcal{E}_{\mathrm{T}}^{2}$ is defined by
(Def. 8) Median $\triangle(A, B, C) \cap$ Median $\triangle(B, C, A)=\{i t\}$ and Median $\triangle(B, C, A) \cap$ Median $\triangle(C, A, B)=\{i t\}$ and Median $\triangle(C, A, B) \cap$ Median $\triangle(A, B, C)=$ $\{i t\}$.

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# Altitude, Orthocenter of a Triangle and Triangulation 

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#### Abstract

Summary. We introduce the altitudes of a triangle (the cevians perpendicular to the opposite sides). Using the generalized Ceva's Theorem, we prove the existence and uniqueness of the orthocenter of a triangle [7]. Finally, we formalize in Mizar [1] some formulas [2] to calculate distance using triangulation.


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## 1. Preliminaries

From now on $n$ denotes a natural number, $i$ denotes an integer, $r, s, t$ denote real numbers, $A_{1}, B_{1}, C_{1}, D_{1}$ denote points of $\mathcal{E}_{\mathrm{T}}^{n}, L_{1}, L_{2}$ denote elements of Lines $\left(\mathcal{R}^{n}\right)$, and $A, B, C$ denote points of $\mathcal{E}_{\mathrm{T}}^{2}$.

Now we state the propositions:
(1) If $0<i \cdot r<r$, then $i=1$.
(2) Let us consider an integer $i$. If $\frac{-3}{2}<i<\frac{1}{2}$, then $i=0$ or $i=-1$.
(3) Suppose $r$ is not zero and $s$ is not zero and $t$ is not zero. Then $\left(\frac{-r}{-s}\right) \cdot$ $\left(\frac{-t}{-r}\right) \cdot\left(\frac{-s}{-t}\right)=1$.
(4) If $0<r<2 \cdot \pi$, then $\sin \left(\frac{r}{2}\right) \neq 0$. The theorem is a consequence of (1).
(5) If $-2 \cdot \pi<r<0$, then $\sin \left(\frac{r}{2}\right) \neq 0$. The theorem is a consequence of (4).
(6) $\tan (2 \cdot \pi-r)=-\tan r$.
(7) If $A_{1} \in \operatorname{Line}\left(B_{1}, C_{1}\right)$ and $A_{1} \neq C_{1}$, then $\operatorname{Line}\left(B_{1}, C_{1}\right)=\operatorname{Line}\left(A_{1}, C_{1}\right)$.
(8) If $A_{1} \neq C_{1}$ and $A_{1} \in \operatorname{Line}\left(B_{1}, C_{1}\right)$, then $B_{1} \in \operatorname{Line}\left(A_{1}, C_{1}\right)$.
(9) Suppose $A_{1} \neq B_{1}$ and $A_{1} \neq C_{1}$ and $\left|\left(A_{1}-B_{1}, A_{1}-C_{1}\right)\right|=0$ and $L_{1}=\operatorname{Line}\left(A_{1}, B_{1}\right)$ and $L_{2}=\operatorname{Line}\left(A_{1}, C_{1}\right)$. Then $L_{1} \perp L_{2}$.
(10) If $B_{1} \neq C_{1}$ and $\left|\left(B_{1}-A_{1}, B_{1}-C_{1}\right)\right|=0$, then $A_{1} \neq C_{1}$.
(11) $\left|\left(A_{1}-B_{1}, A_{1}-C_{1}\right)\right|=\left|\left(B_{1}-A_{1}, C_{1}-A_{1}\right)\right|$.
(12) Suppose $B_{1} \neq C_{1}$ and $r=-\left(\frac{\left|\left(B_{1}, C_{1}\right)\right|-\left|\left(C_{1}, C_{1}\right)\right|-\left|\left(A_{1}, B_{1}\right)\right|+\left|\left(A_{1}, C_{1}\right)\right|}{\left|\left(B_{1}-C_{1}, B_{1}-C_{1}\right)\right|}\right)$ and $D_{1}=r \cdot B_{1}+(1-r) \cdot C_{1}$. Then $\left|\left(D_{1}-A_{1}, D_{1}-C_{1}\right)\right|=0$.
(13) If $A_{1} \neq B_{1}$ and $C_{1}=r \cdot A_{1}+(1-r) \cdot B_{1}$ and $C_{1}=B_{1}$, then $r=0$.
(14) (i) $\left|\left(B_{1}, C_{1}\right)\right|-\left|\left(C_{1}, C_{1}\right)\right|-\left|\left(A_{1}, B_{1}\right)\right|+\left|\left(A_{1}, C_{1}\right)\right|=\left|\left(C_{1}-A_{1}, B_{1}-C_{1}\right)\right|$, and
(ii) $\left|\left(B_{1}-C_{1}, B_{1}-C_{1}\right)\right|+\left|\left(C_{1}-A_{1}, B_{1}-C_{1}\right)\right|=\left|\left(B_{1}-C_{1}, B_{1}-A_{1}\right)\right|$.
(15) $\left|\left(A_{1}-B_{1}, A_{1}-C_{1}\right)\right|=-\left|\left(A_{1}-B_{1}, C_{1}-A_{1}\right)\right|$.
(16) $\left|\left(B_{1}-A_{1}, C_{1}-A_{1}\right)\right|=\left|\left(A_{1}-B_{1}, A_{1}-C_{1}\right)\right|$.
(17) $\left|\left(B_{1}-A_{1}, C_{1}-A_{1}\right)\right|=-\left|\left(B_{1}-A_{1}, A_{1}-C_{1}\right)\right|$. The theorem is a consequence of (16) and (15).
(18) Suppose $B_{1} \neq C_{1}$ and $C_{1} \neq A_{1}$ and $A_{1} \neq B_{1}$ and $\left|\left(C_{1}-A_{1}, B_{1}-C_{1}\right)\right|$ is not zero and $\left|\left(B_{1}-C_{1}, A_{1}-B_{1}\right)\right|$ is not zero and $\mid\left(C_{1}-A_{1}, A_{1}-\right.$ $\left.B_{1}\right) \mid$ is not zero and $r=-\left(\frac{\left|\left(B_{1}, C_{1}\right)\right|-\left|\left(C_{1}, C_{1}\right)\right|-\left|\left(A_{1}, B_{1}\right)\right|+\left|\left(A_{1}, C_{1}\right)\right|}{\left|\left(B_{1}-C_{1}, B_{1}-C_{1}\right)\right|}\right)$ and $s=$ $-\left(\frac{\left|\left(C_{1}, A_{1}\right)\right|-\left|\left(A_{1}, A_{1}\right)\right|-\left|\left(B_{1}, C_{1}\right)\right|+\left|\left(B_{1}, A_{1}\right)\right|}{\left|\left(C_{1}-A_{1}, C_{1}-A_{1}\right)\right|}\right)$ and
$t=-\left(\frac{\left|\left(A_{1}, B_{1}\right)\right|-\left|\left(B_{1}, B_{1}\right)\right|-\left|\left(C_{1}, A_{1}\right)\right|+\left|\left(C_{1}, B_{1}\right)\right|}{\left|\left(A_{1}-B_{1}, A_{1}-B_{1}\right)\right|}\right)$.Then $\frac{\left(\frac{\left(\frac{r}{1-r}\right) \cdot s}{1-s}\right) \cdot t}{1-t}=1$. The theorem is a consequence of (14), (15), and (3).
(19) If $C_{1}=r \cdot A_{1}+(1-r) \cdot B_{1}$ and $r=1$, then $C_{1}=A_{1}$.
(20) If $C_{1}=r \cdot A_{1}+(1-r) \cdot B_{1}$ and $r=0$, then $C_{1}=B_{1}$.
(21) If $\left|\left(B_{1}-C_{1}, B_{1}-C_{1}\right)\right|=-\left|\left(C_{1}-A_{1}, B_{1}-C_{1}\right)\right|$, then $\mid\left(B_{1}-C_{1}, A_{1}-\right.$ $\left.B_{1}\right) \mid=0$. The theorem is a consequence of (15).
(22) Suppose $B_{1} \neq C_{1}$ and $r=-\left(\frac{\left|\left(B_{1}, C_{1}\right)\right|-\left|\left(C_{1}, C_{1}\right)\right|-\left|\left(A_{1}, B_{1}\right)\right|+\left|\left(A_{1}, C_{1}\right)\right|}{\left|\left(B_{1}-C_{1}, B_{1}-C_{1}\right)\right|}\right)$ and $r=1$. Then $\left|\left(B_{1}-C_{1}, A_{1}-B_{1}\right)\right|=0$. The theorem is a consequence of (14) and (21).
(23) If $A \neq B$ and $A \neq C$, then $|A-B|+|A-C| \neq 0$.
(24) If $A, B, C$ form a triangle, then $A \notin \operatorname{Line}(B, C)$.
(25) If $A \neq B$ and $B \neq C$ and $|(B-A, B-C)|=0$, then $\measuredangle(A, B, C)=\frac{\pi}{2}$ or $\measuredangle(A, B, C)=\left(\frac{3}{2}\right) \cdot \pi$.
(26) If $A, B, C$ form a triangle, then $\sin \left(\frac{\measuredangle(A, B, C)}{2}\right)>0$.
(27) If $\measuredangle(B, A, C) \neq \measuredangle(C, B, A)$, then $\sin \left(\frac{\measuredangle(B, A, C)-\measuredangle(C, B, A)}{2}\right) \neq 0$. The theorem is a consequence of (5) and (4).
(28) If $A, B, C$ form a triangle, then $\sin \measuredangle(A, B, C) \neq 0$.

Let us assume that $A, C, B$ form a triangle and $\measuredangle(A, C, B)<\pi$. Now we state the propositions:
(29) $\measuredangle(A, C, B)=\pi-(\measuredangle(C, B, A)+\measuredangle(B, A, C))$.
(30) $\measuredangle(B, A, C)+\measuredangle(C, B, A)=\pi-\measuredangle(A, C, B)$. The theorem is a consequence of (29).
Let us assume that $A, B, C$ form a triangle. Now we state the propositions:

$$
\begin{align*}
& \text { (31) } \measuredangle(B, A, C)-\measuredangle(C, B, A) \neq \pi .  \tag{31}\\
& \text { (32) } \measuredangle(B, A, C)-\measuredangle(C, B, A) \neq-\pi .
\end{align*}
$$

Let us assume that $A, B, C$ form a triangle. Now we state the propositions:

$$
\begin{equation*}
(-2) \cdot \pi<\measuredangle(B, A, C)-\measuredangle(C, B, A)<2 \cdot \pi \tag{33}
\end{equation*}
$$

(34) $-\pi<\frac{\measuredangle(B, A, C)-\measuredangle(C, B, A)}{2}<\pi$. The theorem is a consequence of (33).

Let us assume that $A, B, C$ form a triangle and $\measuredangle(B, A, C)<\pi$. Now we state the propositions:
(35) $-\pi<\measuredangle(B, A, C)-\measuredangle(C, B, A)<\pi$.
(36) $-\left(\frac{\pi}{2}\right)<\frac{\measuredangle(B, A, C)-\measuredangle(C, B, A)}{2}<\frac{\pi}{2}$. The theorem is a consequence of (35).

## 2. Orthocenter

From now on $D$ denotes a point of $\mathcal{E}_{\mathrm{T}}^{2}$ and $a, b, c, d$ denote real numbers.
Let $A, B, C$ be points of $\mathcal{E}_{T}^{2}$. Assume $B \neq C$. The functor Altit $\triangle(A, B, C)$ yielding an element of $\operatorname{Lines}\left(\mathcal{R}^{2}\right)$ is defined by
(Def. 1) there exist elements $L_{1}, L_{2}$ of $\operatorname{Lines}\left(\mathcal{R}^{2}\right)$ such that it $=L_{1}$ and $L_{2}=$ Line $(B, C)$ and $A \in L_{1}$ and $L_{1} \perp L_{2}$.
Let us assume that $B \neq C$. Now we state the propositions:
(37) $A \in$ Altit $\triangle(A, B, C)$.
(38) Altit $\triangle(A, B, C)$ is a line.
$(39) \quad$ Altit $\triangle(A, B, C)=$ Altit $\triangle(A, C, B)$.
Now we state the propositions:
(40) If $B \neq C$ and $D \in$ Altit $\triangle(A, B, C)$, then Altit $\triangle(D, B, C)=$ Altit $\triangle(A, B, C)$.
(41) If $B \neq C$ and $D \in \operatorname{Line}(B, C)$ and $D \neq C$, then Altit $\triangle(A, B, C)=$ Altit $\triangle(A, D, C)$. The theorem is a consequence of (7).
Let $A, B, C$ be points of $\mathcal{E}_{\mathrm{T}}^{2}$. Assume $B \neq C$. The functor FootAltit $\triangle(A, B, C)$ yielding a point of $\mathcal{E}_{\mathrm{T}}^{2}$ is defined by
(Def. 2) there exists a point $P$ of $\mathcal{E}_{\mathrm{T}}^{2}$ such that $i t=P$ and Altit $\triangle(A, B, C) \cap$ Line $(B, C)=\{P\}$.
Let us assume that $B \neq C$. Now we state the propositions:
(42) FootAltit $\triangle(A, B, C)=$ FootAltit $\triangle(A, C, B)$. The theorem is a consequence of (39).
(43) (i) FootAltit $\triangle(A, B, C) \in \operatorname{Line}(B, C)$, and
(ii) FootAltit $\triangle(A, B, C) \in$ Altit $\triangle(A, B, C)$.

Now we state the propositions:
(44) If $B \neq C$ and $A \notin \operatorname{Line}(B, C)$, then Altit $\triangle(A, B, C)=$ Line $(A$, Foot Altit $\triangle(A, B, C))$. The theorem is a consequence of (43).
If $B \neq C$ and $A \in \operatorname{Line}(B, C)$, then FootAltit $\triangle(A, B, C)=A$.
(46) If $B \neq C$ and FootAltit $\triangle(A, B, C)=A$, then $A \in \operatorname{Line}(B, C)$.

Let us assume that $B \neq C$. Now we state the propositions:
(47) $\mid(A-$ FootAltit $\triangle(A, B, C), B-C) \mid=0$. The theorem is a consequence of (44) and (45).
(48) $\mid(A$-FootAltit $\triangle(A, B, C), B$-FootAltit $\triangle(A, B, C)) \mid=0$. The theorem is a consequence of $(43),(44)$, and (45).
(49) $\mid(A$-FootAltit $\triangle(A, B, C), C$-FootAltit $\triangle(A, B, C)) \mid=0$. The theorem is a consequence of (42) and (48).
Now we state the propositions:
(50) If $B \neq C$ and $B=$ FootAltit $\triangle(A, B, C)$, then $|(B-A, B-C)|=0$. The theorem is a consequence of (49), (11), and (43).
(51) If $B \neq C$ and $D \in \operatorname{Line}(B, C)$ and $D \neq C$, then FootAltit $\triangle(A, B, C)=$ FootAltit $\triangle(A, D, C)$. The theorem is a consequence of (7) and (41).
(52) If $B \neq C$ and $|(B-A, B-C)|=0$, then $B=$ FootAltit $\triangle(A, B, C)$. The theorem is a consequence of (9) and (45).
(53) If $B \neq C$ and $B \neq A$ and $\measuredangle(A, B, C)=\frac{\pi}{2}$, then FootAltit $\triangle(A, B, C)=$ $B$. The theorem is a consequence of (11) and (52).
(54) If $A, B, C$ form a triangle, then $A \neq$ FootAltit $\triangle(A, B, C)$. The theorem is a consequence of (43).
(55) If $A, B, C$ form a triangle and $|(B-A, B-C)| \neq 0$, then FootAltit $\triangle(A, B, C), B, A$ form a triangle.
Proof: Set $p=$ FootAltit $\triangle(A, B, C)$. Consider $P$ being a point of $\mathcal{E}_{\mathrm{T}}^{2}$ such that FootAltit $\triangle(A, B, C)=P$ and Altit $\triangle(A, B, C) \cap \operatorname{Line}(B, C)=\{P\}$. Consider $L_{1}, L_{2}$ being elements of $\operatorname{Lines}\left(\mathcal{R}^{2}\right)$ such that Altit $\triangle(A, B, C)=$ $L_{1}$ and $L_{2}=\operatorname{Line}(B, C)$ and $A \in L_{1}$ and $L_{1} \perp L_{2} . P \neq B . p \neq A . p, B, A$ are mutually different. $P \in \operatorname{Line}(B, C) . B, C \in \operatorname{Line}(B, P) . \measuredangle(p, B, A) \neq \pi$
by [11, (11)], [12, (12)], (50), (8). $\measuredangle(B, A, p) \neq \pi$ by [11, (11)], [12, (12)]. $\measuredangle(A, p, B) \neq \pi$ by [11, (11)], [12, (12)], (8), (54).
Let $A, B, C$ be points of $\mathcal{E}_{\mathrm{T}}^{2}$. Assume $B \neq C$. The functor $\mid$ Altit $\triangle(A, B, C) \mid$ yielding a real number is defined by the term
(Def. 3) $\mid A$ - FootAltit $\triangle(A, B, C) \mid$.
Let us assume that $B \neq C$. Now we state the propositions:
(56) $0 \leqslant \mid$ Altit $\triangle(A, B, C) \mid$.
(57) $\mid$ Altit $\triangle(A, B, C)|=|$ Altit $\triangle(A, C, B) \mid$. The theorem is a consequence of (42).
Now we state the propositions:
(58) If $B \neq C$ and $|(B-A, B-C)|=0$, then $\mid$ FootAltit $\triangle(A, B, C)-A \mid=$ $|A-B|$. The theorem is a consequence of (52).
(59) Suppose $B \neq C$ and $r=-\left(\frac{|(B, C)|-|(C, C)|-|(A, B)|+|(A, C)|}{|(B-C, B-C)|}\right)$ and $D=r \cdot B+$ $(1-r) \cdot C$ and $D \neq C$. Then $D=$ FootAltit $\triangle(A, B, C)$.
Proof: $|(D-A, D-C)|=0 . D=$ FootAltit $\triangle(A, D, C) . D \in \operatorname{Line}(B, C)$ by [6, (4)].
(60) Suppose $B \neq C$ and $r=-\left(\frac{|(B, C)|-|(C, C)|-|(A, B)|+|(A, C)|}{|(B-C, B-C)|}\right)$ and $D=r \cdot B+$ $(1-r) \cdot C$ and $D=C$. Then $C=$ FootAltit $\triangle(A, B, C)$. The theorem is a consequence of (13), (14), (15), (52), and (42).
(61) Suppose $A, B, C$ form a triangle and $|(C-A, B-C)|$ is not zero and $|(B-C, A-B)|$ is not zero and $|(C-A, A-B)|$ is not zero. Then Line $(A$, FootAltit $\triangle(A, B, C))$, Line $(C$, FootAltit $\triangle(C, A, B))$,
Line $(B$, FootAltit $\triangle(B, C, A))$ are concurrent. The theorem is a consequence of (60), (17), (47), (59), (18), and (22).
(62) If $A, B, C$ form a triangle and $|(C-A, B-C)|$ is zero, then

FootAltit $\triangle(A, B, C)=C$ and FootAltit $\triangle(B, C, A)=C$. The theorem is a consequence of (15), (52), and (42).
(63) Suppose $A, B, C$ form a triangle and $C \in$ Altit $\triangle(A, B, C)$ and $C \in$ Altit $\triangle(B, C, A)$. Then Altit $\triangle(A, B, C) \cap$ Altit $\triangle(B, C, A)$ is a point. Proof: Consider $L_{1}, L_{2}$ being elements of $\operatorname{Lines}\left(\mathcal{R}^{2}\right)$ such that Altit $\triangle(A, B, C)=L_{1}$ and $L_{2}=\operatorname{Line}(B, C)$ and $A \in L_{1}$ and $L_{1} \perp L_{2}$. Consider $L_{3}, L_{4}$ being elements of $\operatorname{Lines}\left(\mathcal{R}^{2}\right)$ such that Altit $\triangle(B, C, A)=$ $L_{3}$ and $L_{4}=\operatorname{Line}(C, A)$ and $B \in L_{3}$ and $L_{3} \perp L_{4} . L_{1} \nVdash L_{3}$ by [9, (41)], [6, (16)], [8, (108)], [12, (13)]. $L_{1}$ is not a point and $L_{3}$ is not a point.
(64) Suppose $B, C, A$ form a triangle and $C \in$ Altit $\triangle(B, C, A)$ and $C \in$ Altit $\triangle(C, A, B)$. Then Altit $\triangle(B, C, A) \cap$ Altit $\triangle(C, A, B)$ is a point. Proof: Consider $L_{1}, L_{2}$ being elements of $\operatorname{Lines}\left(\mathcal{R}^{2}\right)$ such that

Altit $\triangle(B, C, A)=L_{1}$ and $L_{2}=\operatorname{Line}(C, A)$ and $B \in L_{1}$ and $L_{1} \perp L_{2}$. Consider $L_{3}, L_{4}$ being elements of Lines $\left(\mathcal{R}^{2}\right)$ such that Altit $\triangle(C, A, B)=$ $L_{3}$ and $L_{4}=\operatorname{Line}(A, B)$ and $C \in L_{3}$ and $L_{3} \perp L_{4} . L_{1} \nVdash L_{3}$ by [8, (71), (111)], [6, (16)], [9, (41)]. $L_{1}$ is not a point and $L_{3}$ is not a point.
(65) Suppose $C, A, B$ form a triangle and $C \in$ Altit $\triangle(C, A, B)$ and $C \in$ Altit $\triangle(A, B, C)$. Then Altit $\triangle(C, A, B) \cap$ Altit $\triangle(A, B, C)$ is a point.
Proof: Consider $L_{1}, L_{2}$ being elements of Lines $\left(\mathcal{R}^{2}\right)$ such that
Altit $\triangle(C, A, B)=L_{1}$ and $L_{2}=\operatorname{Line}(A, B)$ and $C \in L_{1}$ and $L_{1} \perp L_{2}$. Consider $L_{3}, L_{4}$ being elements of Lines $\left(\mathcal{R}^{2}\right)$ such that Altit $\triangle(A, B, C)=$ $L_{3}$ and $L_{4}=\operatorname{Line}(B, C)$ and $A \in L_{3}$ and $L_{3} \perp L_{4} . L_{1} \nVdash L_{3}$ by [8, (71), (111)], [6, (16)], [9, (41)]. $L_{1}$ is not a point and $L_{3}$ is not a point.
(66) Suppose $A, B, C$ form a triangle and $|(C-A, B-C)|=0$. Then
(i) Altit $\triangle(A, B, C) \cap$ Altit $\triangle(B, C, A)=\{C\}$, and
(ii) Altit $\triangle(B, C, A) \cap$ Altit $\triangle(C, A, B)=\{C\}$, and
(iii) Altit $\triangle(C, A, B) \cap$ Altit $\triangle(A, B, C)=\{C\}$.

Proof: $A \notin \operatorname{Line}(B, C)$ and $B \notin \operatorname{Line}(C, A)$. FootAltit $\triangle(A, B, C)=$ $C$ and FootAltit $\triangle(B, C, A)=C$. Altit $\triangle(A, B, C)=\operatorname{Line}(A, C)$ and Altit $\triangle(B, C, A)=\operatorname{Line}(B, C) . C \in$ Altit $\triangle(C, A, B)$. Altit $\triangle(A, B, C) \cap$ Altit $\triangle(B, C, A)=\{C\}$ by [6, (22)], (63). Altit $\triangle(B, C, A) \cap$ Altit $\triangle(C, A, B)$ $=\{C\}$ by [12, (15)], (37), (64), [6, (22)]. Altit $\triangle(C, A, B) \cap \operatorname{Altit} \triangle(A, B, C)$ $=\{C\}$ by [12, (15)], (37), (65), [6, (22)].
(67) Suppose $A, B, C$ form a triangle. Then there exists a point $P$ of $\mathcal{E}_{\mathrm{T}}^{2}$ such that
(i) Altit $\triangle(A, B, C) \cap$ Altit $\triangle(B, C, A)=\{P\}$, and
(ii) Altit $\triangle(B, C, A) \cap$ Altit $\triangle(C, A, B)=\{P\}$, and
(iii) Altit $\triangle(C, A, B) \cap$ Altit $\triangle(A, B, C)=\{P\}$.

The theorem is a consequence of (66), (61), (24), (44), and (38).
Let $A, B, C$ be points of $\mathcal{E}_{\mathrm{T}}^{2}$. Assume $A, B, C$ form a triangle. The functor Orthocenter $\triangle(A, B, C)$ yielding a point of $\mathcal{E}_{\mathrm{T}}^{2}$ is defined by
(Def. 4) Altit $\triangle(A, B, C) \cap$ Altit $\triangle(B, C, A)=\{i t\}$ and Altit $\triangle(B, C, A) \cap$ Altit $\triangle(C, A, B)=\{i t\}$ and Altit $\triangle(C, A, B) \cap$ Altit $\triangle(A, B, C)=\{i t\}$.

## 3. Triangulation

Let us assume that $B \neq A$. Now we state the propositions:
(68) $\quad(\sin \measuredangle(B, A, C)+\sin \measuredangle(C, B, A)) \cdot(|C-B|-|C-A|)=(\sin \measuredangle(B, A, C)-$ $\sin \measuredangle(C, B, A)) \cdot(|C-B|+|C-A|)$.
(69) $\sin \left(\frac{\measuredangle(B, A, C)+\measuredangle(C, B, A)}{2}\right) \cdot \cos \left(\frac{\measuredangle(B, A, C)-\measuredangle(C, B, A)}{2}\right) \cdot(|C-B|-|C-A|)=$ $\sin \left(\frac{\measuredangle(B, A, C)-\measuredangle(C, B, A)}{2}\right) \cdot \cos \left(\frac{\measuredangle(B, A, C)+\measuredangle(C, B, A)}{2}\right) \cdot(|C-B|+|C-A|)$. The theorem is a consequence of (68).
Now we state the proposition:
(70) Suppose $A, B, C$ form a triangle and $\measuredangle(B, A, C)-\measuredangle(C, B, A) \neq \pi$ and $\measuredangle(B, A, C)-\measuredangle(C, B, A) \neq-\pi$. Then $\cos \left(\frac{\measuredangle(B, A, C)-\measuredangle(C, B, A)}{2}\right) \neq 0$. The theorem is a consequence of (2).
Let us assume that $A, C, B$ form a triangle and $\measuredangle(A, C, B)<\pi$. Now we state the propositions:
(71) $\tan \left(\frac{\measuredangle(B, A, C)-\measuredangle(C, B, A)}{2}\right)=\cot \left(\frac{\measuredangle(A, C, B)}{2}\right) \cdot\left(\frac{|C-B|-|C-A|}{|C-B|+|C-A|}\right)$.

Proof: $\measuredangle(B, A, C)-\measuredangle(C, B, A) \neq \pi$ and $\measuredangle(B, A, C)-\measuredangle(C, B, A) \neq-\pi$.
Set $\alpha=\frac{\measuredangle(B, A, C)-\measuredangle(C, B, A)}{2}$. Set $\beta=\frac{\measuredangle(B, A, C)+\measuredangle(C, B, A)}{2} . \measuredangle(A, C, B)=\pi-$
$(\measuredangle(C, B, A)+\measuredangle(B, A, C))$. Set $\alpha_{1}=\frac{\measuredangle(A, C, B)}{2} . \sin \alpha_{1} \neq 0 .|C-B|+|C-A| \neq$ 0 by [11, (42)]. $\sin \beta \cdot \cos \alpha \cdot(|C-B|-|C-A|)=\sin \alpha \cdot \cos \beta \cdot(|C-B|+$ $|C-A|) \cdot(|C-B|-|C-A|) \cdot \cos \alpha_{1} \cdot 1=(|C-B|+|C-A|) \cdot \sin \alpha_{1} \cdot\left(\frac{\sin \alpha}{\cos \alpha}\right)$.
(72) $\frac{\measuredangle(B, A, C)-\measuredangle(C, B, A)}{2}=\arctan \left(\cot \left(\frac{\measuredangle(A, C, B)}{2}\right) \cdot\left(\frac{|C-B|-|C-A|}{|C-B|+|C-A|}\right)\right)$. The theorem is a consequence of (71) and (36).
(73) $\measuredangle(B, A, C)-\measuredangle(C, B, A)=2 \cdot \arctan \left(\cot \left(\frac{\measuredangle(A, C, B)}{2}\right) \cdot\left(\frac{|C-B|-|C-A|}{|C-B|+|C-A|}\right)\right)$. The theorem is a consequence of (72).
(i) $\measuredangle(B, A, C)=\arctan \left(\cot \left(\frac{\measuredangle(A, C, B)}{2}\right) \cdot\left(\frac{|C-B|-|C-A|}{|C-B|+|C-A|}\right)\right)+\left(\frac{\pi}{2}\right)-\left(\frac{\measuredangle(A, C, B)}{2}\right)$, and
(ii) $\measuredangle(C, B, A)=\left(\frac{\pi}{2}\right)-\left(\frac{\measuredangle(A, C, B)}{2}\right)-\arctan \left(\cot \left(\frac{\measuredangle(A, C, B)}{2}\right) \cdot\left(\frac{|C-B|-|C-A|}{|C-B|+|C-A|}\right)\right)$. The theorem is a consequence of (73) and (30).
$|B-C|=\frac{|A-B| \cdot \sin \measuredangle(B, A, C)}{\sin (\measuredangle(B, A, C)+\measuredangle(C, B, A))}$.
Proof: $|B-C|=\frac{|A-B| \cdot \sin \measuredangle(B, A, C)}{\sin \measuredangle(A, C, B)}$ by [11, (6), (43)], (28). $\measuredangle(A, C, B)=$ $\pi-(\measuredangle(C, B, A)+\measuredangle(B, A, C))$.
(76) $|A-C|=\frac{|A-B| \cdot \sin \measuredangle(C, B, A)}{\sin (\measuredangle(B, A, C)+\measuredangle(C, B, A))}$.

Proof: $|A-C|=\frac{|A-B| \cdot \sin \measuredangle(C, B, A)}{\sin \measuredangle(A, C, B)}$ by [11, (6)], (28). $\measuredangle(A, C, B)=\pi-$ $(\measuredangle(C, B, A)+\measuredangle(B, A, C))$ by [11, (20)], [10, (47)].
Now we state the propositions:
(77) Suppose $A, C, B$ form a triangle and $\measuredangle(C, A, B)=\frac{\pi}{2}$.

Then $\mid$ Altit $\triangle(C, A, B)|=|A-B| \cdot \tan \measuredangle(A, B, C)$. The theorem is a consequence of (11) and (58).
(78) Suppose $A, B, C$ form a triangle and $\measuredangle(C, A, B)=\left(\frac{3}{2}\right) \cdot \pi$.

Then $\mid$ Altit $\triangle(C, A, B)|=|A-B| \cdot \tan \measuredangle(C, B, A)$. The theorem is a consequence of (11) and (58).
(79) Suppose $A, C, B$ form a triangle and $|(A-C, A-B)|=0$. Then $\mid$ Altit $\triangle(C, A, B)|=|A-B| \cdot| \tan \measuredangle(A, B, C) \mid$. The theorem is a consequence of (11), (77), (56), (6), and (78).
(80) Suppose $B \neq C$ and FootAltit $\triangle(A, B, C), B, A$ form a triangle. Then
(i) $|A-B| \cdot \sin \measuredangle(A, B$, FootAltit $\triangle(A, B, C))=\mid$ FootAltit $\triangle(A, B, C)-$ $A \mid$, or
(ii) $|A-B| \cdot(-\sin \measuredangle(A, B$, FootAltit $\triangle(A, B, C)))=$ $\mid$ FootAltit $\triangle(A, B, C)-A \mid$.
The theorem is a consequence of (48).
(81) Suppose $A, B, C$ form a triangle and $|(B-A, B-C)| \neq 0$. Then
(i) $|A-B| \cdot \sin \measuredangle(A, B$, FootAltit $\triangle(A, B, C))=\mid$ FootAltit $\triangle(A, B, C)-$ $A \mid$, or
(ii) $|A-B| \cdot(-\sin \measuredangle(A, B$, FootAltit $\triangle(A, B, C)))=$ $\mid$ Foot Altit $\triangle(A, B, C)-A \mid$.
The theorem is a consequence of (80) and (55).
(82) Suppose $A, C, B$ form a triangle and $\measuredangle(A, C, B)<\pi$ and $\mid(A-C, A-$ $B) \mid \neq 0$. Then $\mid$ Altit $\triangle(C, A, B)|=|A-B| \cdot|\left(\frac{\sin \measuredangle(C, B, A)}{\sin (\measuredangle(B, A, C)+\measuredangle(C, B, A))}\right)$. $\sin \measuredangle(C, A$, FootAltit $\triangle(C, A, B)) \mid$. The theorem is a consequence of $(76)$, (55), and (80).
(83) Suppose $0<\measuredangle(B, A, D)<\pi$ and $0<\measuredangle(D, A, C)<\pi$ and $D, A, C$ are mutually different and $B, A, D$ are mutually different. Then $\measuredangle(A, C, D)+$ $\measuredangle(D, B, A)=2 \cdot \pi-(\measuredangle(B, A, C)+\measuredangle(A, D, B)+\measuredangle(C, D, A))$.
Proof: $\measuredangle(B, A, D)+\measuredangle(D, A, C)=\measuredangle(B, A, C)$ by [5, (2)], [11, (4)]. $\measuredangle(A, C$, $D)=\pi-(\measuredangle(C, D, A)+\measuredangle(D, A, C))$ by [10, (47)]. $\measuredangle(D, B, A)=\pi-$ $(\measuredangle(A, D, B)+\measuredangle(B, A, D))$ by [10, (47)].
(84) Suppose $A, C, B$ form a triangle and $\measuredangle(A, C, B)<\pi$ and $A, D, B$ form a triangle and $\measuredangle(A, D, B)<\pi$ and $a=\measuredangle(C, B, A)$ and $b=\measuredangle(B, A, C)$ and $c=\measuredangle(D, B, A)$ and $d=\measuredangle(C, A, D)$. Then $|D-C|^{2}=|A-B|^{2}$. $\left(\left(\frac{\sin a}{\sin (a+b)}\right)^{2}+\left(\frac{\sin c}{\sin (b+d+c)}\right)^{2}-2 \cdot\left(\frac{\sin a}{\sin (b+a)}\right) \cdot\left(\frac{\sin c}{\sin (b+d+c)}\right) \cdot \cos d\right)$.
Proof: Set $e=b+d . \sin (e+c)=\sin (\measuredangle(B, A, D)+\measuredangle(D, B, A))$ by [14, (79)].
(85) Suppose $\sin (2 \cdot s) \cdot \cos d=\cos (2 \cdot t)$. Then $(r \cdot \cos s)^{2}+(r \cdot \sin s)^{2}-2$. $(r \cdot \cos s) \cdot(r \cdot \sin s) \cdot \cos d=2 \cdot r^{2} \cdot(\sin t)^{2}$.
(86) Let us consider real numbers $R, \vartheta$. Suppose $D \neq C$ and $0 \leqslant R$ and $A$, $C, B$ form a triangle and $\measuredangle(A, C, B)<\pi$ and $A, D, B$ form a triangle
and $\measuredangle(A, D, B)<\pi$ and $a=\measuredangle(C, B, A)$ and $b=\measuredangle(B, A, C)$ and $c=$ $\measuredangle(D, B, A)$ and $d=\measuredangle(C, A, D)$ and $R \cdot \cos s=\frac{\sin a}{\sin (a+b)}$ and $R \cdot \sin s=$ $\frac{\sin c}{\sin (b+d+c)}$ and $0<\vartheta<\pi$ and $\sin (2 \cdot s) \cdot \cos d=\cos (2 \cdot \vartheta)$. Then $|D-C|=$ $|A-B| \cdot \sqrt{2} \cdot R \cdot \sin \vartheta$.
Proof: $|D-C|^{2}=|A-B|^{2} \cdot\left((R \cdot \cos s)^{2}+(R \cdot \sin s)^{2}-2 \cdot(R \cdot \cos s) \cdot(R\right.$. $\sin s) \cdot \cos d) \cdot|D-C| \neq-|A-B| \cdot \sqrt{2} \cdot R \cdot \sin \vartheta$ by [13, (25)], [11, (42)].
(87) Suppose $A, C, B$ form a triangle and $\measuredangle(A, C, B)<\pi$ and $D, A, C$ form a triangle and $\measuredangle(A, D, C)=\frac{\pi}{2}$. Then $|D-C|=\left(\frac{|A-B| \cdot \sin \measuredangle(C, B, A)}{\sin (\measuredangle(B, A, C)+\measuredangle(C, B, A))}\right)$. $\sin \measuredangle(C, A, D)$. The theorem is a consequence of (76).
(88) Suppose $B, C, A$ form a triangle and $\measuredangle(B, C, A)<\pi$ and $D, C, A$ form a triangle and $\measuredangle(C, D, A)=\frac{\pi}{2}$. Then $|D-C|=\left(\frac{|A-B| \cdot \sin \measuredangle(A, B, C)}{\sin (\measuredangle(A, B, C)+\measuredangle(C, A, B))}\right)$. $\sin \measuredangle(D, A, C)$. The theorem is a consequence of (75).
(89) Suppose $A, C, B$ form a triangle and $\measuredangle(A, C, B)<\pi$ and $D, A, C$ form a triangle and $\measuredangle(A, D, C)=\frac{\pi}{2}$ and $A \in \mathcal{L}(B, D)$ and $A \neq D$. Then $|D-C|=$ $\left(\frac{|A-B| \cdot \sin \measuredangle(C, B, A)}{\sin (\measuredangle(C, A, D)-\measuredangle(C, B, A))}\right) \cdot \sin \measuredangle(C, A, D)$. The theorem is a consequence of (87).
(90) Suppose $B, C, A$ form a triangle and $\measuredangle(B, C, A)<\pi$ and $D, C, A$ form a triangle and $\measuredangle(C, D, A)=\frac{\pi}{2}$ and $A \in \mathcal{L}(D, B)$ and $A \neq D$. Then $|D-C|=\left(\frac{|A-B| \cdot \sin \measuredangle(A, B, C)}{\sin (\measuredangle(D, A, C)-\measuredangle(A, B, C))}\right) \cdot \sin \measuredangle(D, A, C)$.
Proof: $\sin (\measuredangle(C, A, B)+\measuredangle(A, B, C))=\sin (\measuredangle(D, A, C)-\measuredangle(A, B, C))$ by [4, (1)], [3, (8)].

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# Divisible $\mathbb{Z}$-modules 

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#### Abstract

Summary. In this article, we formalize the definition of divisible $\mathbb{Z}$-module and its properties in the Mizar system [3]. We formally prove that any non-trivial divisible $\mathbb{Z}$-modules are not finitely-generated. We introduce a divisible $\mathbb{Z}$-module, equivalent to a vector space of a torsion-free $\mathbb{Z}$-module with a coefficient ring $\mathbb{Q}$. $\mathbb{Z}$-modules are important for lattice problems, LLL (Lenstra, Lenstra and Lovász) base reduction algorithm [15, cryptographic systems with lattices 16] and coding theory 8 .


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## 1. Divisible Module

Let $a, b$ be elements of $\mathbb{F}_{\mathbb{Q}}$ and $x, y$ be rational numbers. We identify $x+y$ with $a+b$. We identify $x \cdot y$ with $a \cdot b$. Let $V$ be a $\mathbb{Z}$-module and $v$ be a vector of $V$. We say that $v$ is divisible if and only if
(Def. 1) for every element $a$ of $\mathbb{Z}^{\mathrm{R}}$ such that $a \neq 0_{\mathbb{Z}^{\mathrm{R}}}$ there exists a vector $u$ of $V$ such that $a \cdot u=v$.

Let us observe that $0_{V}$ is divisible and there exists a vector of $V$ which is divisible.

Now we state the propositions:
(1) Let us consider a $\mathbb{Z}$-module $V$, and divisible vectors $v, u$ of $V$. Then $v+u$ is divisible.
(2) Let us consider a $\mathbb{Z}$-module $V$, and a divisible vector $v$ of $V$. Then $-v$ is divisible.
Proof: For every element $a$ of $\mathbb{Z}^{\mathrm{R}}$ such that $a \neq 0_{\mathbb{Z}^{\mathrm{R}}}$ there exists a vector $w$ of $V$ such that $-v=a \cdot w$ by [9, (6)]. $\square$
(3) Let us consider a $\mathbb{Z}$-module $V$, a divisible vector $v$ of $V$, and an element $i$ of $\mathbb{Z}^{\mathrm{R}}$. Then $i \cdot v$ is divisible.
Let $V$ be a $\mathbb{Z}$-module. We say that $V$ is divisible if and only if
(Def. 2) every vector of $V$ is divisible.
Observe that $\mathbf{0}_{V}$ is divisible and $\mathbb{Z}$-module $\mathbb{Q}$ is divisible and there exists a $\mathbb{Z}$-module which is divisible.

Let $V$ be a $\mathbb{Z}$-module. Let us note that there exists a submodule of $V$ which is divisible and there exists a divisible $\mathbb{Z}$-module which is non finitely generated.

Now we state the propositions:
(4) (The left integer multiplication of $\left.\mathbb{F}_{\mathbb{Q}}\right) \upharpoonright(\mathbb{Z} \times \mathbb{Z})=$ the left integer multiplication of $\mathbb{Z}^{\mathrm{R}}$.
Proof: Set $a=\left(\right.$ the left integer multiplication of $\left.\mathbb{F}_{\mathbb{Q}}\right) \upharpoonright(\mathbb{Z} \times \mathbb{Z})$. For every object $z$ such that $z \in \operatorname{dom} a$ holds $a(z)=$ (the left integer multiplication of $\left.\mathbb{Z}^{\mathrm{R}}\right)(z)$ by [5, (49)], [13, (15)], [12, (14)].
(5) <the carrier of $\mathbb{Z}^{\mathrm{R}}$, the addition of $\mathbb{Z}^{\mathrm{R}}$, the zero of $\mathbb{Z}^{\mathrm{R}}$, the left integer multiplication of $\left.\mathbb{Z}^{\mathrm{R}}\right\rangle$ is a submodule of $\mathbb{Z}$-module $\mathbb{Q}$. The theorem is a consequence of (4).
(6) Let us consider a divisible $\mathbb{Z}$-module $V$, and a submodule $W$ of $V$. Then $\mathbb{Z}$-ModuleQuot $(V, W)$ is divisible.
Let us note that there exists a divisible $\mathbb{Z}$-module which is non trivial.
Now we state the proposition:
(7) Let us consider a $\mathbb{Z}$-module $V$. Then $V$ is divisible if and only if $\Omega_{V}$ is divisible.

Let us consider a $\mathbb{Z}$-module $V$ and a vector $v$ of $V$. Now we state the propositions:
(8) If $v$ is not torsion, then $\operatorname{Lin}(\{v\})$ is not divisible.
(9) If $v$ is torsion and $v \neq 0_{V}$, then $\operatorname{Lin}(\{v\})$ is not divisible.

Let $V$ be a non trivial $\mathbb{Z}$-module and $v$ be a non zero vector of $V$. Observe that $\operatorname{Lin}(\{v\})$ is non divisible and there exists a submodule of $V$ which is non divisible.

Now we state the propositions:
(10) Every non trivial, finitely generated, torsion-free $\mathbb{Z}$-module is not divisible.

Proof: Consider $I$ being a finite subset of $V$ such that $I$ is a basis of $V$. Consider $v$ being an object such that $v \in I . v$ is not divisible by [9, (92)], [12, (19)], [19, (15)], [9, (9)].
(11) Let us consider a non trivial, finitely generated, torsion $\mathbb{Z}$-module $V$. Then there exists an element $i$ of $\mathbb{Z}^{\mathrm{R}}$ such that
(i) $i \neq 0$, and
(ii) for every vector $v$ of $V, i \cdot v=0_{V}$.

Proof: Define $\mathcal{P}$ [natural number] $\equiv$ for every finite subset $I$ of $V$ such that $\overline{\bar{I}}=\$_{1}$ there exists an element $i$ of $\mathbb{Z}^{\mathrm{R}}$ such that $i \neq 0$ and for every vector $v$ of $V$ such that $v \in \operatorname{Lin}(I)$ holds $i \cdot v=0_{V} . \mathcal{P}[0]$ by [10, (67)], [9, (1)]. For every natural number $n$ such that $\mathcal{P}[n]$ holds $\mathcal{P}[n+1]$ by [7, (40)], [10, (72)], [1, (44)], [7, (31)]. For every natural number $n, \mathcal{P}[n]$ from [2, Sch. 2]. Consider $I$ being a finite subset of $V$ such that $\operatorname{Lin}(I)=$ the vector space structure of $V$. Consider $i$ being an element of $\mathbb{Z}^{\mathrm{R}}$ such that $i \neq 0$ and for every vector $v$ of $V$ such that $v \in \operatorname{Lin}(I)$ holds $i \cdot v=0_{V}$. For every vector $v$ of $V, i \cdot v=0_{V}$.
(12) Let us consider a non trivial, finitely generated, torsion $\mathbb{Z}$-module $V$, and an element $i$ of $\mathbb{Z}^{\mathrm{R}}$. Suppose $i \neq 0$ and for every vector $v$ of $V, i \cdot v=0_{V}$. Then $V$ is not divisible.
(13) Every non trivial, finitely generated, torsion $\mathbb{Z}$-module is not divisible. The theorem is a consequence of (11) and (12).
One can verify that there exists a non trivial, finitely generated, torsion $\mathbb{Z}$-module which is non divisible.

Now we state the proposition:
(14) Every non trivial, finitely generated $\mathbb{Z}$-module is not divisible. The theorem is a consequence of (13), (6), and (10).
Let us note that every non trivial, divisible $\mathbb{Z}$-module is non finitely generated.

Let $V$ be a non trivial, non divisible $\mathbb{Z}$-module. One can verify that there exists a non zero vector of $V$ which is non divisible.

Let $V$ be a non trivial, finite rank, free $\mathbb{Z}$-module. Observe that rank $V$ is non zero.

Now we state the propositions:
(15) Let us consider a non trivial, free $\mathbb{Z}$-module $V$, a non zero vector $v$ of $V$, and a basis $I$ of $V$. Then there exists a linear combination $L$ of $I$ and there exists a vector $u$ of $V$ such that $v=\sum L$ and $u \in I$ and $L(u) \neq 0$. Proof: Consider $L$ being a linear combination of $I$ such that $v=\sum L$. The support of $L \neq \emptyset$ by [10, (23)]. Consider $u_{1}$ being an object such that
$u_{1} \in$ the support of $L$. Consider $u$ being a vector of $V$ such that $u=u_{1}$ and $L(u) \neq 0$.
(16) Let us consider a non trivial, free $\mathbb{Z}$-module $V$. Then every non zero vector of $V$ is not divisible. The theorem is a consequence of (15).
Let us observe that every non trivial, free $\mathbb{Z}$-module is non divisible.
Let us consider a non trivial, free $\mathbb{Z}$-module $V$ and a non zero vector $v$ of $V$.
Now we state the propositions:
(17) There exists an element $a$ of $\mathbb{Z}^{\mathrm{R}}$ such that
(i) $a \in \mathbb{N}$, and
(ii) for every element $b$ of $\mathbb{Z}^{\mathrm{R}}$ and for every vector $u$ of $V$ such that $b>a$ holds $v \neq b \cdot u$.
Proof: Set $I=$ the basis of $V$. Consider $L$ being a linear combination of $I, w$ being a vector of $V$ such that $v=\sum L$ and $w \in I$ and $L(w) \neq 0$. Reconsider $a=|L(w)|$ as an element of $\mathbb{Z}^{\mathrm{R}}$. For every element $b$ of $\mathbb{Z}^{\mathrm{R}}$ and for every vector $u$ of $V$ such that $b>a$ holds $v \neq b \cdot u$ by [10, (64), (31), (53)], [11, (3)].
(18) There exists an element $a$ of $\mathbb{Z}^{\mathrm{R}}$ and there exists a vector $u$ of $V$ such that $a \in \mathbb{N}$ and $a \neq 0$ and $v=a \cdot u$ and for every element $b$ of $\mathbb{Z}^{\mathrm{R}}$ and for every vector $w$ of $V$ such that $b>a$ holds $v \neq b \cdot w$.
Proof: Define $\mathcal{P}$ [natural number] $\equiv$ there exists a vector $u$ of $V$ and there exists an element $k$ of $\mathbb{Z}^{\mathrm{R}}$ such that $k=\$_{1}$ and $v=k \cdot u$. Consider $a$ being an element of $\mathbb{Z}^{\mathrm{R}}$ such that $a \in \mathbb{N}$ and for every element $b$ of $\mathbb{Z}^{\mathrm{R}}$ and for every vector $u$ of $V$ such that $b>a$ holds $v \neq b \cdot u$. There exists a natural number $k$ such that $\mathcal{P}[k]$. Consider $a_{0}$ being a natural number such that $\mathcal{P}\left[a_{0}\right]$ and for every natural number $n$ such that $\mathcal{P}[n]$ holds $n \leqslant a_{0}$ from [2, Sch. 6]. Reconsider $a=a_{0}$ as an element of $\mathbb{Z}^{\mathrm{R}}$. Consider $u$ being a vector of $V$ such that $v=a \cdot u . a \neq 0$ by [9, (1)]. For every element $b$ of $\mathbb{Z}^{\mathrm{R}}$ and for every vector $w$ of $V$ such that $b>a$ holds $v \neq b \cdot w$ by [18, (3)].

## 2. Divisible Module for Torsion-free $\mathbb{Z}$-module

Let $V$ be a torsion-free $\mathbb{Z}$-module. The functor $\operatorname{Embedding}(V)$ yielding a strict $\mathbb{Z}$-module is defined by
(Def. 3) the carrier of $i t=\operatorname{rng} \operatorname{MorphsZQ}(V)$ and the zero of $i t=\operatorname{zeroCoset}(V)$ and the addition of $i t=\operatorname{addCoset}(V) \upharpoonright \operatorname{rng} \operatorname{MorphsZQ}(V)$ and the left multiplication of $i t=\operatorname{lmult} \operatorname{Coset}(V) \upharpoonright(\mathbb{Z} \times \operatorname{rng} \operatorname{MorphsZQ}(V))$.
Let us consider a torsion-free $\mathbb{Z}$-module $V$. Now we state the propositions:
(19) (i) every vector of $\operatorname{Embedding}(V)$ is a vector of $\mathbb{Z}-\operatorname{MQVectSp}(V)$, and
(ii) $0_{\text {Embedding }(V)}=0_{\mathbb{Z}-\operatorname{MQVectSp}(V)}$, and
(iii) for every vectors $x, y$ of $\operatorname{Embedding}(V)$ and for every vectors $v, w$ of $\mathbb{Z}-\operatorname{MQVectSp}(V)$ such that $x=v$ and $y=w$ holds $x+y=v+w$, and
(iv) for every element $i$ of $\mathbb{Z}^{\mathrm{R}}$ and for every element $j$ of $\mathbb{F}_{\mathbb{Q}}$ and for every vector $x$ of $\operatorname{Embedding}(V)$ and for every vector $v$ of $\mathbb{Z}-\operatorname{MQVectSp}(V)$ such that $i=j$ and $x=v$ holds $i \cdot x=j \cdot v$.
Proof: Set $Z=\mathbb{Z}-\operatorname{MQVectSp}(V)$. Set $E=\operatorname{Embedding}(V)$. For every vectors $x, y$ of $E$ and for every vectors $v, w$ of $Z$ such that $x=v$ and $y=w$ holds $x+y=v+w$ by [5, (49)]. For every element $i$ of $\mathbb{Z}^{\mathrm{R}}$ and for every element $j$ of $\mathbb{F}_{\mathbb{Q}}$ and for every vector $x$ of $E$ and for every vector $v$ of $Z$ such that $i=j$ and $x=v$ holds $i \cdot x=j \cdot v$ by [5, (49)].
(20) (i) for every vectors $v, w$ of $\mathbb{Z}-\operatorname{MQVectSp}(V)$ such that $v, w \in \operatorname{Embedding}(V)$ holds $v+w \in \operatorname{Embedding}(V)$, and
(ii) for every element $j$ of $\mathbb{F}_{\mathbb{Q}}$ and for every vector $v$ of $\mathbb{Z}$ - $\operatorname{MQVectSp}(V)$ such that $j \in \mathbb{Z}$ and $v \in \operatorname{Embedding}(V)$ holds $j \cdot v \in \operatorname{Embedding}(V)$. The theorem is a consequence of (19).
(21) There exists a linear transformation $T$ from $V$ to $\operatorname{Embedding}(V)$ such that
(i) $T$ is bijective, and
(ii) $T=\operatorname{MorphsZQ}(V)$, and
(iii) for every vector $v$ of $V, T(v)=[\langle v, 1\rangle]_{\operatorname{EQRZM}(V)}$.

The theorem is a consequence of (19).
Now we state the proposition:
(22) Let us consider a torsion-free $\mathbb{Z}$-module $V$, and a vector $v_{1}$ of $\operatorname{Embedding}(V)$. Then there exists a vector $v$ of $V$ such that $(\operatorname{MorphsZQ}(V))(v)=v_{1}$. The theorem is a consequence of (21).
Let $V$ be a torsion-free $\mathbb{Z}$-module. The functor $\operatorname{DivisibleMod}(V)$ yielding a strict $\mathbb{Z}$-module is defined by
(Def. 4) the carrier of $i t=$ Classes $\operatorname{EQRZM}(V)$ and the zero of $i t=\operatorname{zeroCoset}(V)$ and the addition of $i t=\operatorname{addCoset}(V)$ and the left multiplication of $i t=$ $\operatorname{lmultCoset}(V) \upharpoonright(\mathbb{Z} \times \operatorname{Classes} \operatorname{EQRZM}(V))$.
Now we state the proposition:
(23) Let us consider a torsion-free $\mathbb{Z}$-module $V$, a vector $v$ of $\operatorname{DivisibleMod}(V)$, and an element $a$ of $\mathbb{Z}^{\mathrm{R}}$. Suppose $a \neq 0$. Then there exists a vector $u$ of DivisibleMod $(V)$ such that $a \cdot u=v$.

Proof: For every vector $v$ of $\operatorname{DivisibleMod}(V)$ and for every element $a$ of $\mathbb{Z}^{\mathrm{R}}$ such that $a \neq 0$ there exists a vector $u$ of $\operatorname{DivisibleMod}(V)$ such that $a \cdot u=v$ by [5, (49)], [7, (87)].
Let $V$ be a torsion-free $\mathbb{Z}$-module. Let us observe that $\operatorname{DivisibleMod}(V)$ is divisible.

Now we state the proposition:
(24) Let us consider a torsion-free $\mathbb{Z}$-module $V$. Then $\operatorname{Embedding}(V)$ is a submodule of DivisibleMod $(V)$.
Proof: Set $E=\operatorname{Embedding}(V)$. Set $D=\operatorname{DivisibleMod}(V)$. For every object $x$ such that $x \in$ the carrier of $E$ holds $x \in$ the carrier of $D$ by [6, (11), (5)]. The left multiplication of $E=$ (the left multiplication of $D) \upharpoonright\left(\left(\right.\right.$ the carrier of $\left.\left.\mathbb{Z}^{\mathrm{R}}\right) \times \operatorname{rng} \operatorname{MorphsZQ}(V)\right)$ by [20, (74)], [7, (96)].
Let $V$ be a finitely generated, torsion-free $\mathbb{Z}$-module. One can check that Embedding $(V)$ is finitely generated.

Let $V$ be a non trivial, torsion-free $\mathbb{Z}$-module. Observe that Embedding $(V)$ is non trivial.

Let $G$ be a field, $V$ be a vector space over $G, W$ be a subset of $V$, and $a$ be an element of $G$. The functor $a \cdot W$ yielding a subset of $V$ is defined by the term
(Def. 5) $\quad\{a \cdot u$, where $u$ is a vector of $V: u \in W\}$.
Let $V$ be a torsion-free $\mathbb{Z}$-module and $r$ be an element of $\mathbb{F}_{\mathbb{Q}}$. The functor Embedding $(r, V)$ yielding a strict $\mathbb{Z}$-module is defined by
(Def. 6) the carrier of it $=r \cdot r$ rng $\operatorname{MorphsZQ}(V)$ and the zero of $i t=\operatorname{zeroCoset}(V)$ and the addition of $i t=\operatorname{addCoset}(V) \upharpoonright(r \cdot$ rng $\operatorname{MorphsZQ}(V))$ and the left multiplication of $i t=$
$\operatorname{lmultCoset}(V) \upharpoonright\left(\left(\right.\right.$ the carrier of $\left.\left.\mathbb{Z}^{\mathrm{R}}\right) \times(r \cdot \operatorname{rng} \operatorname{MorphsZQ}(V))\right)$.
Let us consider a torsion-free $\mathbb{Z}$-module $V$ and an element $r$ of $\mathbb{F}_{\mathbb{Q}}$. Now we state the propositions:
(25) (i) every vector of $\operatorname{Embedding}(r, V)$ is a vector of $\mathbb{Z}$ - $\operatorname{MQVectSp}(V)$, and
(ii) $0_{\text {Embedding }(r, V)}=0_{\mathbb{Z}-\operatorname{MQVectSp}(V)}$, and
(iii) for every vectors $x, y$ of $\operatorname{Embed} \operatorname{ding}(r, V)$ and for every vectors $v, w$ of $\mathbb{Z}$-MQVectSp $(V)$ such that $x=v$ and $y=w$ holds $x+y=v+w$, and
(iv) for every element $i$ of $\mathbb{Z}^{\mathrm{R}}$ and for every element $j$ of $\mathbb{F}_{\mathbb{Q}}$ and for every vector $x$ of Embedding $(r, V)$ and for every vector $v$ of $\mathbb{Z}$ - $\operatorname{MQVectSp}(V)$ such that $i=j$ and $x=v$ holds $i \cdot x=j \cdot v$.
Proof: Set $Z=\mathbb{Z}$-MQVectSp $(V)$. Set $E=\operatorname{Embedding}(r, V)$. For every vectors $x, y$ of $E$ and for every vectors $v, w$ of $Z$ such that $x=v$ and
$y=w$ holds $x+y=v+w$ by [5, (49)]. For every element $i$ of $\mathbb{Z}^{\mathrm{R}}$ and for every element $j$ of $\mathbb{F}_{\mathbb{Q}}$ and for every vector $x$ of $E$ and for every vector $v$ of $Z$ such that $i=j$ and $x=v$ holds $i \cdot x=j \cdot v$ by [5, (49)].
(i) for every vectors $v, w$ of $\mathbb{Z}-\operatorname{MQVectSp}(V)$ such that $v, w \in \operatorname{Embedding}(r, V)$ holds $v+w \in \operatorname{Embedding}(r, V)$, and
(ii) for every element $j$ of $\mathbb{F}_{\mathbb{Q}}$ and for every vector $v$ of $\mathbb{Z}-\operatorname{MQVectSp}(V)$ such that $j \in \mathbb{Z}$ and $v \in \operatorname{Embedding}(r, V)$ holds $j \cdot v \in \operatorname{Embedding}(r, V)$. The theorem is a consequence of (25).
(27) Suppose $r \neq 0_{\mathbb{F}_{\mathbb{Q}}}$. Then there exists a linear transformation $T$ from Embedding $(V)$ to Embedding $(r, V)$ such that
(i) for every element $v$ of $\mathbb{Z}$ - $\operatorname{MQVectSp}(V)$ such that $v \in \operatorname{Embedding}(V)$ holds $T(v)=r \cdot v$, and
(ii) $T$ is bijective.

Proof: Set $Z=\mathbb{Z}$-MQVectSp $(V)$. Define $\mathcal{F}$ (vector of $Z)=r \cdot \$_{1}$. Consider $T$ being a function from the carrier of $Z$ into the carrier of $Z$ such that for every element $x$ of the carrier of $Z, T(x)=\mathcal{F}(x)$ from [6, Sch. 4]. Set $T_{0}=T \upharpoonright$ (the carrier of Embedding $\left.(V)\right)$. For every object $y, y \in \operatorname{rng} T_{0}$ iff $y \in$ the carrier of Embedding $(r, V)$ by [5, (49)]. $T_{0}$ is additive by (19), (20), [5, (49)], (25). For every element $x$ of $\operatorname{Embedding}(V)$ and for every element $i$ of $\mathbb{Z}^{\mathrm{R}}, T_{0}(i \cdot x)=i \cdot T_{0}(x)$ by (19), (20), [5, (49)], (25). For every element $v$ of $\mathbb{Z}-\operatorname{MQVectSp}(V)$ such that $v \in \operatorname{Embedding}(V)$ holds $T_{0}(v)=r \cdot v$ by [5, (49)]. For every objects $x_{1}, x_{2}$ such that $x_{1}, x_{2} \in$ the carrier of Embedding $(V)$ and $T_{0}\left(x_{1}\right)=T_{0}\left(x_{2}\right)$ holds $x_{1}=x_{2}$ by [14, (20)].
Now we state the propositions:
(28) Let us consider a torsion-free $\mathbb{Z}$-module $V$, and a vector $v$ of $V$. Then $[\langle v, 1\rangle]_{\operatorname{EQRZM}(V)} \in \operatorname{Embedding}(V)$.
(29) Let us consider a torsion-free $\mathbb{Z}$-module $V$, and a vector $v$ of DivisibleMod $(V)$. Then there exists an element $a$ of $\mathbb{Z}^{\mathrm{R}}$ such that
(i) $a \neq 0$, and
(ii) $a \cdot v \in \operatorname{Embedding}(V)$.

The theorem is a consequence of (28).
Let $V$ be a torsion-free $\mathbb{Z}$-module. One can check that $\operatorname{DivisibleMod}(V)$ is torsion-free and Embedding $(V)$ is torsion-free.

Let $V$ be a free $\mathbb{Z}$-module. Let us note that $\operatorname{Embedding}(V)$ is free.
Let us consider a torsion-free $\mathbb{Z}$-module $V$. Now we state the propositions:
(i) every vector of $\mathbb{Z}$ - $\operatorname{MQVectSp}(V)$ is a vector of $\operatorname{DivisibleMod}(V)$, and
(ii) every vector of $\operatorname{DivisibleMod}(V)$ is a vector of $\mathbb{Z}$ - $\operatorname{MQVectSp}(V)$, and
(iii) $0_{\text {DivisibleMod }(V)}=0_{\mathbb{Z}-\operatorname{MQVectSp}(V)}$.
(31) (i) for every vectors $x, y$ of $\operatorname{DivisibleMod}(V)$ and for every vectors $v$, $u$ of $\mathbb{Z}-\operatorname{MQVectSp}(V)$ such that $x=v$ and $y=u$ holds $x+y=v+u$, and
(ii) for every vector $z$ of $\operatorname{DivisibleMod}(V)$ and for every vector $w$ of $\mathbb{Z}$-MQVectSp $(V)$ and for every element $a$ of $\mathbb{Z}^{\mathrm{R}}$ and for every element $a_{1}$ of $\mathbb{F}_{\mathbb{Q}}$ such that $z=w$ and $a=a_{1}$ holds $a \cdot z=a_{1} \cdot w$, and
(iii) for every vector $z$ of $\operatorname{DivisibleMod}(V)$ and for every vector $w$ of $\mathbb{Z}-\operatorname{MQVectSp}(V)$ and for every element $a_{1}$ of $\mathbb{F}_{\mathbb{Q}}$ and for every element $a$ of $\mathbb{Z}^{\mathrm{R}}$ such that $a \neq 0$ and $a_{1}=a$ and $a \cdot z=a_{1} \cdot w$ holds $z=w$, and
(iv) for every vector $x$ of $\operatorname{DivisibleMod}(V)$ and for every vector $v$ of $\mathbb{Z}-\operatorname{MQVectSp}(V)$ and for every element $r$ of $\mathbb{F}_{\mathbb{Q}}$ and for every elements $m, n$ of $\mathbb{Z}^{\mathrm{R}}$ and for every integers $m_{1}, n_{1}$ such that $m=m_{1}$ and $n=n_{1}$ and $x=v$ and $r \neq 0_{\mathbb{F}_{\mathbb{Q}}}$ and $n \neq 0$ and $r=\frac{m_{1}}{n_{1}}$ there exists a vector $y$ of $\operatorname{DivisibleMod}(V)$ such that $x=n \cdot y$ and $r \cdot v=m \cdot y$.
Proof: For every vector $z$ of $\operatorname{DivisibleMod}(V)$ and for every vector $w$ of $\mathbb{Z}$-MQVectSp $(V)$ and for every element $a$ of $\mathbb{Z}^{\mathrm{R}}$ and for every element $a_{1}$ of $\mathbb{F}_{\mathbb{Q}}$ such that $z=w$ and $a=a_{1}$ holds $a \cdot z=a_{1} \cdot w$ by [5, (49)], [7, (87)]. For every vector $z$ of $\operatorname{DivisibleMod}(V)$ and for every vector $w$ of $\mathbb{Z}-\mathrm{MQVectSp}(V)$ and for every element $a_{1}$ of $\mathbb{F}_{\mathbb{Q}}$ and for every element $a$ of $\mathbb{Z}^{\mathrm{R}}$ such that $a \neq 0$ and $a_{1}=a$ and $a \cdot z=a_{1} \cdot w$ holds $z=w$ by (30), [9, (8)], [19, (15), (21)]. For every vector $x$ of DivisibleMod( $V$ ) and for every vector $v$ of $\mathbb{Z}$-MQVectSp $(V)$ and for every element $r$ of $\mathbb{F}_{\mathbb{Q}}$ and for every elements $m, n$ of $\mathbb{Z}^{\mathrm{R}}$ and for every integers $m_{1}, n_{1}$ such that $m=m_{1}$ and $n=n_{1}$ and $x=v$ and $r \neq 0_{\mathbb{F}_{\mathbb{Q}}}$ and $n \neq 0$ and $r=\frac{m_{1}}{n_{1}}$ there exists a vector $y$ of DivisibleMod $(V)$ such that $x=n \cdot y$ and $r \cdot v=m \cdot y$.
Now we state the proposition:
(32) Let us consider a torsion-free $\mathbb{Z}$-module $V$, and an element $r$ of $\mathbb{F}_{\mathbb{Q}}$. Then Embedding $(r, V)$ is a submodule of $\operatorname{DivisibleMod}(V)$. The theorem is a consequence of (25) and (30).
Let $V$ be a finitely generated, torsion-free $\mathbb{Z}$-module and $r$ be an element of $\mathbb{F}_{\mathbb{Q}}$. Observe that Embedding $(r, V)$ is finitely generated.

Let $V$ be a non trivial, torsion-free $\mathbb{Z}$-module and $r$ be a non zero element of $\mathbb{F}_{\mathbb{Q}}$. One can verify that Embedding $(r, V)$ is non trivial.

Let $V$ be a torsion-free $\mathbb{Z}$-module and $r$ be an element of $\mathbb{F}_{\mathbb{Q}}$. Observe that Embedding $(r, V)$ is torsion-free.

Let $V$ be a free $\mathbb{Z}$-module and $r$ be a non zero element of $\mathbb{F}_{\mathbb{Q}}$. One can check that Embedding $(r, V)$ is free.

Now we state the propositions:
(33) Let us consider a non trivial, free $\mathbb{Z}$-module $V$, and a vector $v$ of DivisibleMod $(V)$. Then there exists an element $a$ of $\mathbb{Z}^{\mathrm{R}}$ such that
(i) $a \in \mathbb{N}$, and
(ii) $a \neq 0$, and
(iii) $a \cdot v \in \operatorname{Embedding}(V)$, and
(iv) for every element $b$ of $\mathbb{Z}^{\mathrm{R}}$ such that $b \in \mathbb{N}$ and $b<a$ and $b \neq 0$ holds $b \cdot v \notin \operatorname{Embedding}(V)$.

Proof: Consider $a_{1}$ being an element of $\mathbb{Z}^{\mathrm{R}}$ such that $a_{1} \neq 0$ and $a_{1} \cdot v \in$ $\operatorname{Embedding}(V) .\left|a_{1}\right| \cdot v \in \operatorname{Embedding}(V)$ by (24), [9, (16), (30)]. Define $\mathcal{P}$ [natural number] $\equiv$ there exists an element $n$ of $\mathbb{Z}^{\mathrm{R}}$ such that $n=\$_{1}$ and $n \in \mathbb{N}$ and $n \neq 0$ and $n \cdot v \in \operatorname{Embedding}(V)$. There exists a natural number $k$ such that $\mathcal{P}[k]$ and for every natural number $n$ such that $\mathcal{P}[n]$ holds $k \leqslant n$ from [2, Sch. 5]. Consider $a_{0}$ being a natural number such that $\mathcal{P}\left[a_{0}\right]$ and for every natural number $b_{0}$ such that $\mathcal{P}\left[b_{0}\right]$ holds $a_{0} \leqslant b_{0}$.
(34) Let us consider a finite rank, free $\mathbb{Z}$-module $V$. Then $\operatorname{rank} \operatorname{Embedding}(V)=$ rank $V$. The theorem is a consequence of (21).
Let us consider a finite rank, free $\mathbb{Z}$-module $V$ and a non zero element $r$ of $\mathbb{F}_{\mathbb{Q}}$. Now we state the propositions:
(35) $\operatorname{rank} \operatorname{Embedding}(r, V)=\operatorname{rank} \operatorname{Embedding}(V)$. The theorem is a consequence of (27).
(36) $\operatorname{rank} \operatorname{Embedding}(r, V)=\operatorname{rank} V$. The theorem is a consequence of (35) and (34).
Observe that every non trivial, torsion-free $\mathbb{Z}$-module is infinite.
Now we state the propositions:
(37) Let us consider a $\mathbb{Z}$-module $V$. Then there exists a subset $A$ of $V$ such that
(i) $A$ is linearly independent, and
(ii) for every vector $v$ of $V$, there exists an element $a$ of $\mathbb{Z}^{\mathrm{R}}$ such that $a \in \mathbb{N}$ and $a>0$ and $a \cdot v \in \operatorname{Lin}(A)$.

Proof: Consider $A$ being a subset of $V$ such that $\emptyset \subseteq A$ and $A$ is linearly independent and for every vector $v$ of $V$, there exists an element $a_{1}$ of $\mathbb{Z}^{\mathrm{R}}$ such that $a_{1} \neq 0$ and $a_{1} \cdot v \in \operatorname{Lin}(A)$. For every vector $v$ of $V$, there exists
an element $a$ of $\mathbb{Z}^{\mathrm{R}}$ such that $a \in \mathbb{N}$ and $a>0$ and $a \cdot v \in \operatorname{Lin}(A)$ by [17, (2)], 44, (46)], [18, (3)], [9, (16), (38)].
(38) Let us consider a non trivial, torsion-free $\mathbb{Z}$-module $V$, a non zero vector $v$ of $V$, a subset $A$ of $V$, and an element $a$ of $\mathbb{Z}^{\mathrm{R}}$. Suppose $a \in \mathbb{N}$ and $A$ is linearly independent and $a>0$ and $a \cdot v \in \operatorname{Lin}(A)$. Then there exists a linear combination $L$ of $A$ and there exists a vector $u$ of $V$ such that $a \cdot v=\sum L$ and $u \in A$ and $L(u) \neq 0$.
Proof: Consider $L$ being a linear combination of $A$ such that $a \cdot v=\sum L$. The support of $L \neq \emptyset$ by [10, (23)]. Consider $u_{1}$ being an object such that $u_{1} \in$ the support of $L$. Consider $u$ being a vector of $V$ such that $u=u_{1}$ and $L(u) \neq 0$.
(39) Let us consider a torsion-free $\mathbb{Z}$-module $V$, a non zero integer $i$, and non zero elements $r_{1}, r_{2}$ of $\mathbb{F}_{\mathbb{Q}}$. Suppose $r_{2}=\frac{r_{1}}{i}$. Then Embedding $\left(r_{1}, V\right)$ is a submodule of Embedding $\left(r_{2}, V\right)$.
Proof: For every vector $x$ of $\operatorname{DivisibleMod}(V)$ such that $x \in \operatorname{Embedding}\left(r_{1}\right.$, $V)$ holds $x \in \operatorname{Embedding}\left(r_{2}, V\right)$ by (27), [6, (11)], (19), [6, (5)]. Embedding $\left(r_{1}, V\right)$ is a submodule of $\operatorname{DivisibleMod}(V)$ and $\operatorname{Embedding}\left(r_{2}, V\right)$ is a submodule of DivisibleMod( $V$ ).
(40) Let us consider a finite rank, free $\mathbb{Z}$-module $V$, and a submodule $Z$ of DivisibleMod $(V)$. Then $Z$ is finitely generated if and only if there exists a non zero element $r$ of $\mathbb{F}_{\mathbb{Q}}$ such that $Z$ is a submodule of Embedding $(r, V)$. The theorem is a consequence of (32), (29), (19), (27), (31), and (39).

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# Lattice of $\mathbb{Z}$-module 

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Summary. In this article, we formalize the definition of lattice of $\mathbb{Z}$-module and its properties in the Mizar system 5. We formally prove that scalar products in lattices are bilinear forms over the field of real numbers $\mathbb{R}$. We also formalize the definitions of positive definite and integral lattices and their properties. Lattice of $\mathbb{Z}$-module is necessary for lattice problems, LLL (Lenstra, Lenstra and Lovász) base reduction algorithm [14, and cryptographic systems with lattices 15 and coding theory [9.

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## 1. Definition of Lattices of $\mathbb{Z}$-module

Now we state the proposition:
(1) Let us consider non empty sets $D, E$, natural numbers $n, m$, and a matrix $M$ over $D$ of dimension $n \times m$. Suppose for every natural numbers $i, j$ such that $\langle i, j\rangle \in$ the indices of $M$ holds $M_{i, j} \in E$. Then $M$ is a matrix over $E$ of dimension $n \times m$.

Let $a, b$ be elements of $\mathbb{F}_{\mathbb{Q}}$ and $x, y$ be rational numbers. We identify $x+y$ with $a+b$ and $x \cdot y$ with $a \cdot b$. Let $F$ be a 1 -sorted structure. We consider structures of $\mathbb{Z}$-lattice over $F$ which extend vector space structures over $F$ and are systems

〈a carrier, an addition, a zero, a left multiplication,

## a scalar product $\rangle$

where the carrier is a set, the addition is a binary operation on the carrier, the zero is an element of the carrier, the left multiplication is a function from (the carrier of $F) \times($ the carrier $)$ into the carrier, the scalar product is a function from (the carrier) $\times$ (the carrier) into the carrier of $\mathbb{R}_{F}$.

Note that there exists a structure of $\mathbb{Z}$-lattice over $F$ which is strict and non empty.

Let $D$ be a non empty set, $Z$ be an element of $D, a$ be a binary operation on $D, m$ be a function from (the carrier of $F) \times D$ into $D$, and $s$ be a function from $D \times D$ into the carrier of $\mathbb{R}_{\mathrm{F}}$. One can check that $\langle D, a, Z, m, s\rangle$ is non empty.

Let $X$ be a non empty structure of $\mathbb{Z}$-lattice over $\mathbb{Z}^{\mathrm{R}}$ and $x, y$ be vectors of $X$. The functor $\langle x, y\rangle$ yielding an element of $\mathbb{R}_{\mathrm{F}}$ is defined by the term
(Def. 1) (the scalar product of $X)(\langle x, y\rangle)$.
Let $x$ be a vector of $X$. The functor $\|x\|$ yielding an element of $\mathbb{R}_{F}$ is defined by the term
(Def. 2) $\langle x, x\rangle$.
Let $X$ be a non empty structure of $\mathbb{Z}$-lattice over $\mathbb{Z}^{\mathrm{R}}$. We say that $X$ is $\mathbb{Z}$-lattice-like if and only if
(Def. 3) for every vector $x$ of $X$ such that for every vector $y$ of $X,\langle x, y\rangle=0$ holds $x=0_{X}$ and for every vectors $x, y$ of $X,\langle x, y\rangle=\langle y, x\rangle$ and for every vectors $x, y, z$ of $X$ and for every element $a$ of $\mathbb{Z}^{\mathrm{R}},\langle x+y, z\rangle=\langle x, z\rangle+\langle y, z\rangle$ and $\langle a \cdot x, y\rangle=a \cdot\langle x, y\rangle$.
Let $V$ be a $\mathbb{Z}$-module and $s$ be a function from (the carrier of $V) \times($ the carrier of $V$ ) into the carrier of $\mathbb{R}_{F}$. The functor $\operatorname{GenLat}(V, s)$ yielding a non empty structure of $\mathbb{Z}$-lattice over $\mathbb{Z}^{R}$ is defined by the term
(Def. 4) $\left\langle\right.$ the carrier of $V$, the addition of $V, 0_{V}$, the left multiplication of $\left.V, s\right\rangle$.
Let us note that there exists a non empty structure of $\mathbb{Z}$-lattice over $\mathbb{Z}^{R}$ which is vector distributive, scalar distributive, scalar associative, scalar unital, Abelian, add-associative, right zeroed, right complementable, and strict.

Let $V$ be a $\mathbb{Z}$-module and $s$ be a function from (the carrier of $V) \times($ the carrier of $V$ ) into the carrier of $\mathbb{R}_{F}$. One can verify that $\operatorname{GenLat}(V, s)$ is Abelian, addassociative, right zeroed, right complementable, scalar distributive, vector distributive, scalar associative, and scalar unital.

Let us consider a $\mathbb{Z}$-module $V$ and a function $s$ from (the carrier of $V$ ) $\times$ (the carrier of $V$ ) into the carrier of $\mathbb{R}_{\mathrm{F}}$. Now we state the propositions:
(2) $\operatorname{GenLat}(V, s)$ is a submodule of $V$.
(3) $V$ is a submodule of $\operatorname{GenLat}(V, s)$.

Note that there exists an Abelian, add-associative, right zeroed, right complementable, scalar distributive, vector distributive, scalar associative, scalar unital, non empty structure of $\mathbb{Z}$-lattice over $\mathbb{Z}^{\mathrm{R}}$ which is free.

Let $V$ be a free $\mathbb{Z}$-module and $s$ be a function from (the carrier of $V$ ) $\times$ (the carrier of $V$ ) into the carrier of $\mathbb{R}_{\mathrm{F}}$. Let us observe that $\operatorname{GenLat}(V, s)$ is free and there exists an Abelian, add-associative, right zeroed, right complementable, scalar distributive, vector distributive, scalar associative, scalar unital, non empty structure of $\mathbb{Z}$-lattice over $\mathbb{Z}^{\mathrm{R}}$ which is torsion-free.

Now we state the proposition:
(4) Let us consider a finite rank, free $\mathbb{Z}$-module $V$, and a function $s$ from (the carrier of $V) \times($ the carrier of $V)$ into the carrier of $\mathbb{R}_{\mathrm{F}}$.
Then GenLat $(V, s)$ is finite rank. The theorem is a consequence of (2).
Let us note that there exists a free, Abelian, add-associative, right zeroed, right complementable, scalar distributive, vector distributive, scalar associative, scalar unital, non empty structure of $\mathbb{Z}$-lattice over $\mathbb{Z}^{R}$ which is finite rank.

Let $V$ be a finite rank, free $\mathbb{Z}$-module and $s$ be a function from (the carrier of $V) \times($ the carrier of $V)$ into the carrier of $\mathbb{R}_{\mathrm{F}}$. Let us note that $\operatorname{GenLat}(V, s)$ is finite rank.

Now we state the proposition:
(5) Let us consider a finite rank, free $\mathbb{Z}$-module $V$, and a function $f$ from (the carrier of $\left.\mathbf{0}_{V}\right) \times\left(\right.$ the carrier of $\left.\mathbf{0}_{V}\right)$ into the carrier of $\mathbb{R}_{\mathrm{F}}$. Suppose $f=\left(\right.$ the carrier of $\left.\mathbf{0}_{V}\right) \times\left(\right.$ the carrier of $\left.\mathbf{0}_{V}\right) \longmapsto 0_{\mathbb{R}_{F}}$. Then $\operatorname{GenLat}\left(\mathbf{0}_{V}, f\right)$ is $\mathbb{Z}$-lattice-like.
Proof: Set $X=\operatorname{GenLat}\left(\mathbf{0}_{V}, f\right)$. For every vector $x$ of $X$ such that for every vector $y$ of $X,\langle x, y\rangle=0$ holds $x=0_{X}$ by [10, (26)]. For every vectors $x, y, z$ of $X$ and for every element $a$ of $\mathbb{Z}^{\mathrm{R}},\langle x, y\rangle=\langle y, x\rangle$ and $\langle x+y, z\rangle=\langle x, z\rangle+\langle y, z\rangle$ and $\langle a \cdot x, y\rangle=a \cdot\langle x, y\rangle$ by [16, (7)], [8, (87)].

Note that there exists a non empty structure of $\mathbb{Z}$-lattice over $\mathbb{Z}^{R}$ which is $\mathbb{Z}$-lattice-like and there exists a finite rank, free, Abelian, add-associative, right zeroed, right complementable, scalar distributive, vector distributive, scalar associative, scalar unital, non empty structure of $\mathbb{Z}$-lattice over $\mathbb{Z}^{R}$ which is $\mathbb{Z}$-lattice-like.

There exists a finite rank, free, $\mathbb{Z}$-lattice-like, Abelian, add-associative, right zeroed, right complementable, scalar distributive, vector distributive, scalar associative, scalar unital, non empty structure of $\mathbb{Z}$-lattice over $\mathbb{Z}^{\mathrm{R}}$ which is strict.

A $\mathbb{Z}$-lattice is a finite rank, free, $\mathbb{Z}$-lattice-like, Abelian, add-associative,
right zeroed, right complementable, scalar distributive, vector distributive, scalar associative, scalar unital, non empty structure of $\mathbb{Z}$-lattice over $\mathbb{Z}^{R}$. Now we state the proposition:
(6) Let us consider a non trivial, torsion-free $\mathbb{Z}$-module $V$, a submodule $Z$ of $V$, a non zero vector $v$ of $V$, and a function $f$ from (the carrier of $Z) \times($ the carrier of $Z)$ into the carrier of $\mathbb{R}_{F}$. Suppose $Z=\operatorname{Lin}(\{v\})$ and for every vectors $v_{1}, v_{2}$ of $Z$ and for every elements $a, b$ of $\mathbb{Z}^{\mathrm{R}}$ such that $v_{1}=a \cdot v$ and $v_{2}=b \cdot v$ holds $f\left(v_{1}, v_{2}\right)=a \cdot b$. Then $\operatorname{GenLat}(Z, f)$ is $\mathbb{Z}$-lattice-like.
Proof: Set $L=\operatorname{GenLat}(Z, f) . L$ is $\mathbb{Z}$-lattice-like by [10, (26)], [12, (19)], [10, (1)], [12, (21)].
Observe that there exists a $\mathbb{Z}$-lattice which is non trivial.
Let $V$ be a torsion-free $\mathbb{Z}$-module. Let us observe that $\mathbb{Z}$ - $\operatorname{MQVectSp}(V)$ is scalar distributive, vector distributive, scalar associative, scalar unital, addassociative, right zeroed, right complementable, and Abelian as a non empty vector space structure over $\mathbb{F}_{\mathbb{Q}}$.

Now we state the propositions:
(7) Let us consider a $\mathbb{Z}$-lattice $L$, and vectors $v, u$ of $L$. Then
(i) $\langle v,-u\rangle=-\langle v, u\rangle$, and
(ii) $\langle-v, u\rangle=-\langle v, u\rangle$.
(8) Let us consider a $\mathbb{Z}$-lattice $L$, and vectors $v, u, w$ of $L$. Then $\langle v, u+w\rangle=$ $\langle v, u\rangle+\langle v, w\rangle$.
(9) Let us consider a $\mathbb{Z}$-lattice $L$, vectors $v, u$ of $L$, and an element $a$ of $\mathbb{Z}^{\mathrm{R}}$. Then $\langle v, a \cdot u\rangle=a \cdot\langle v, u\rangle$.
(10) Let us consider a $\mathbb{Z}$-lattice $L$, vectors $v, u$, $w$ of $L$, and elements $a, b$ of $\mathbb{Z}^{\mathrm{R}}$. Then
(i) $\langle a \cdot v+b \cdot u, w\rangle=a \cdot\langle v, w\rangle+b \cdot\langle u, w\rangle$, and
(ii) $\langle v, a \cdot u+b \cdot w\rangle=a \cdot\langle v, u\rangle+b \cdot\langle v, w\rangle$.

The theorem is a consequence of (8) and (9).
(11) Let us consider a $\mathbb{Z}$-lattice $L$, and vectors $v, u, w$ of $L$. Then
(i) $\langle v-u, w\rangle=\langle v, w\rangle-\langle u, w\rangle$, and
(ii) $\langle v, u-w\rangle=\langle v, u\rangle-\langle v, w\rangle$.

The theorem is a consequence of (8) and (9).
(12) Let us consider a $\mathbb{Z}$-lattice $L$, and a vector $v$ of $L$. Then
(i) $\left\langle v, 0_{L}\right\rangle=0$, and
(ii) $\left\langle 0_{L}, v\right\rangle=0$.

The theorem is a consequence of (11).
Let $X$ be a $\mathbb{Z}$-lattice. We say that $X$ is integral if and only if
(Def. 5) for every vectors $v, u$ of $X,\langle v, u\rangle \in \mathbb{Z}$.
Observe that there exists a $\mathbb{Z}$-lattice which is integral.
Let $L$ be an integral $\mathbb{Z}$-lattice and $v, u$ be vectors of $L$. Let us observe that $\langle v, u\rangle$ is integer.

Let $v$ be a vector of $L$. Let us note that $\|v\|$ is integer.
Now we state the propositions:
(13) Let us consider a $\mathbb{Z}$-lattice $L$, a finite subset $I$ of $L$, and a vector $u$ of $L$. Suppose for every vector $v$ of $L$ such that $v \in I$ holds $\langle v, u\rangle \in \mathbb{Z}$. Let us consider a vector $v$ of $L$. If $v \in \operatorname{Lin}(I)$, then $\langle v, u\rangle \in \mathbb{Z}$.
Proof: Define $\mathcal{P}$ [natural number] $\equiv$ for every finite subset $I$ of $L$ such that $\overline{\bar{I}}=\$_{1}$ and for every vector $v$ of $L$ such that $v \in I$ holds $\langle v, u\rangle \in \mathbb{Z}$ for every vector $v$ of $L$ such that $v \in \operatorname{Lin}(I)$ holds $\langle v, u\rangle \in \mathbb{Z} . \mathcal{P}[0]$ by [11, (67)], (12). For every natural number $n$ such that $\mathcal{P}[n]$ holds $\mathcal{P}[n+1]$ by [8, (40)], [11, (72)], [1, (44)], [8, (31)]. For every natural number $n, \mathcal{P}[n]$ from [3, Sch. 2].
(14) Let us consider a $\mathbb{Z}$-lattice $L$, and a basis $I$ of $L$. Suppose for every vectors $v, u$ of $L$ such that $v, u \in I$ holds $\langle v, u\rangle \in \mathbb{Z}$. Let us consider vectors $v, u$ of $L$. Then $\langle v, u\rangle \in \mathbb{Z}$.
Proof: Define $\mathcal{P}$ [natural number] $\equiv$ for every finite subset $I$ of $L$ such that $\overline{\bar{I}}=\$_{1}$ and for every vectors $v, u$ of $L$ such that $v, u \in I$ holds $\langle v, u\rangle \in \mathbb{Z}$ for every vectors $v, u$ of $L$ such that $v, u \in \operatorname{Lin}(I)$ holds $\langle v, u\rangle \in \mathbb{Z}$. $\mathcal{P}[0]$ by [11, (67)], (12). For every natural number $n$ such that $\mathcal{P}[n]$ holds $\mathcal{P}[n+1]$ by [8, (40)], [11, (72)], [1, (44)], [8, (31)]. For every natural number $n, \mathcal{P}[n]$ from [3, Sch. 2].
(15) Let us consider a $\mathbb{Z}$-lattice $L$, and a basis $I$ of $L$. Suppose for every vectors $v, u$ of $L$ such that $v, u \in I$ holds $\langle v, u\rangle \in \mathbb{Z}$. Then $L$ is integral.
Let $X$ be a $\mathbb{Z}$-lattice. We say that $X$ is positive definite if and only if
(Def. 6) for every vector $v$ of $X$ such that $v \neq 0_{X}$ holds $\|v\|>0$.
Let us observe that there exists a $\mathbb{Z}$-lattice which is non trivial, integral, and positive definite.

Let us consider a positive definite $\mathbb{Z}$-lattice $L$ and a vector $v$ of $L$. Now we state the propositions:
(16) $\|v\|=0$ if and only if $v=0_{L}$.
(17) $\|v\| \geqslant 0$. The theorem is a consequence of (12).

Let $X$ be an integral $\mathbb{Z}$-lattice. We say that $X$ is even if and only if (Def. 7) for every vector $v$ of $X,\|v\|$ is even.

One can verify that there exists an integral $\mathbb{Z}$-lattice which is even.
Let $L$ be a $\mathbb{Z}$-lattice. We introduce the notation $\operatorname{dim}(L)$ as a synonym of $\operatorname{rank} L$.

Let $v, u$ be vectors of $L$. We say that $v, u$ are orthogonal if and only if (Def. 8) $\langle v, u\rangle=0$.

Let us note that the predicate is symmetric.
Let us consider a $\mathbb{Z}$-lattice $L$ and vectors $v, u$ of $L$.
Let us assume that $v, u$ are orthogonal. Now we state the propositions:
(18) (i) $v,-u$ are orthogonal, and
(ii) $-v, u$ are orthogonal, and
(iii) $-v,-u$ are orthogonal.

The theorem is a consequence of (7).
(19) $\|v+u\|=\|v\|+\|u\|$. The theorem is a consequence of (8).
(20) $\quad\|v-u\|=\|v\|+\|u\|$. The theorem is a consequence of (11).

Let $L$ be a $\mathbb{Z}$-lattice.
A $\mathbb{Z}$-sublattice of $L$ is a $\mathbb{Z}$-lattice and is defined by
(Def. 9) the carrier of it $\subseteq$ the carrier of $L$ and $0_{i t}=0_{L}$ and the addition of $i t=($ the addition of $L) \upharpoonright($ the carrier of $i t)$ and the left multiplication of $i t=($ the left multiplication of $L) \upharpoonright\left(\left(\right.\right.$ the carrier of $\left.\mathbb{Z}^{\mathrm{R}}\right) \times($ the carrier of $\left.i t)\right)$ and the scalar product of $i t=($ the scalar product of $L) \upharpoonright$ (the carrier of $i t)$.
Now we state the propositions:
(21) Let us consider a $\mathbb{Z}$-lattice $L$. Then every $\mathbb{Z}$-sublattice of $L$ is a submodule of $L$.
(22) Let us consider an object $x$, a $\mathbb{Z}$-lattice $L$, and $\mathbb{Z}$-sublattices $L_{1}, L_{2}$ of $L$. Suppose $x \in L_{1}$ and $L_{1}$ is a $\mathbb{Z}$-sublattice of $L_{2}$. Then $x \in L_{2}$. The theorem is a consequence of (21).
(23) Let us consider an object $x$, a $\mathbb{Z}$-lattice $L$, and a $\mathbb{Z}$-sublattice $L_{1}$ of $L$. If $x \in L_{1}$, then $x \in L$. The theorem is a consequence of (21).
(24) Let us consider a $\mathbb{Z}$-lattice $L$, and a $\mathbb{Z}$-sublattice $L_{1}$ of $L$. Then every vector of $L_{1}$ is a vector of $L$. The theorem is a consequence of (21).
(25) Let us consider a $\mathbb{Z}$-lattice $L$, and $\mathbb{Z}$-sublattices $L_{1}, L_{2}$ of $L$. Then $0_{L_{1}}=$ $0_{L_{2}}$.
(26) Let us consider a $\mathbb{Z}$-lattice $L$, a $\mathbb{Z}$-sublattice $L_{1}$ of $L$, vectors $v_{1}, v_{2}$ of $L$, and vectors $w_{1}, w_{2}$ of $L_{1}$. If $w_{1}=v_{1}$ and $w_{2}=v_{2}$, then $w_{1}+w_{2}=v_{1}+v_{2}$. The theorem is a consequence of (21).
(27) Let us consider a $\mathbb{Z}$-lattice $L$, a $\mathbb{Z}$-sublattice $L_{1}$ of $L$, a vector $v$ of $L$, a vector $w$ of $L_{1}$, and an element $a$ of $\mathbb{Z}^{\mathrm{R}}$. If $w=v$, then $a \cdot w=a \cdot v$. The theorem is a consequence of (21).
(28) Let us consider a $\mathbb{Z}$-lattice $L$, a $\mathbb{Z}$-sublattice $L_{1}$ of $L$, a vector $v$ of $L$, and a vector $w$ of $L_{1}$. If $w=v$, then $-w=-v$. The theorem is a consequence of (21).
(29) Let us consider a $\mathbb{Z}$-lattice $L$, a $\mathbb{Z}$-sublattice $L_{1}$ of $L$, vectors $v_{1}, v_{2}$ of $L$, and vectors $w_{1}, w_{2}$ of $L_{1}$. If $w_{1}=v_{1}$ and $w_{2}=v_{2}$, then $w_{1}-w_{2}=v_{1}-v_{2}$. The theorem is a consequence of (21).
(30) Let us consider a $\mathbb{Z}$-lattice $L$, and a $\mathbb{Z}$-sublattice $L_{1}$ of $L$. Then $0_{L} \in L_{1}$. The theorem is a consequence of (21).
(31) Let us consider a $\mathbb{Z}$-lattice $L$, and $\mathbb{Z}$-sublattices $L_{1}, L_{2}$ of $L$. Then $0_{L_{1}} \in$ $L_{2}$. The theorem is a consequence of (21).
(32) Let us consider a $\mathbb{Z}$-lattice $L$, and a $\mathbb{Z}$-sublattice $L_{1}$ of $L$. Then $0_{L_{1}} \in L$. The theorem is a consequence of (21).
(33) Let us consider a $\mathbb{Z}$-lattice $L$, a $\mathbb{Z}$-sublattice $L_{1}$ of $L$, and vectors $v_{1}, v_{2}$ of $L$. If $v_{1}, v_{2} \in L_{1}$, then $v_{1}+v_{2} \in L_{1}$. The theorem is a consequence of (21).
(34) Let us consider a $\mathbb{Z}$-lattice $L$, a $\mathbb{Z}$-sublattice $L_{1}$ of $L$, a vector $v$ of $L$, and an element $a$ of $\mathbb{Z}^{\mathrm{R}}$. If $v \in L_{1}$, then $a \cdot v \in L_{1}$. The theorem is a consequence of (21).
(35) Let us consider a $\mathbb{Z}$-lattice $L$, a $\mathbb{Z}$-sublattice $L_{1}$ of $L$, and a vector $v$ of $L$. If $v \in L_{1}$, then $-v \in L_{1}$. The theorem is a consequence of (21).
(36) Let us consider a $\mathbb{Z}$-lattice $L$, a $\mathbb{Z}$-sublattice $L_{1}$ of $L$, and vectors $v_{1}, v_{2}$ of $L$. If $v_{1}, v_{2} \in L_{1}$, then $v_{1}-v_{2} \in L_{1}$. The theorem is a consequence of (21).
(37) Let us consider a positive definite $\mathbb{Z}$-lattice $L$, a non empty set $A$, an element $z$ of $A$, a binary operation $a$ on $A$, a function $m$ from (the carrier of $\left.\mathbb{Z}^{\mathrm{R}}\right) \times A$ into $A$, and a function $s$ from $A \times A$ into the carrier of $\mathbb{R}_{\mathrm{F}}$. Suppose $A$ is a linearly closed subset of $L$ and $z=0_{L}$ and $a=$ (the addition of $L) \upharpoonright A$ and $m=($ the left multiplication of $L) \upharpoonright\left(\left(\right.\right.$ the carrier of $\left.\left.\mathbb{Z}^{\mathrm{R}}\right) \times A\right)$ and $s=($ the scalar product of $L) \upharpoonright A$. Then $\langle A, a, z, m, s\rangle$ is a $\mathbb{Z}$-sublattice of $L$.
Proof: Set $L_{1}=\langle A, a, z, m, s\rangle$. Set $V_{1}=\langle A, a, z, m\rangle . L_{1}$ is a submodule of $V_{1} . L_{1}$ is $\mathbb{Z}$-lattice-like by [10, (25)], [7, (49)], [10, (28), (29)].
(38) Let us consider a $\mathbb{Z}$-lattice $L$, a $\mathbb{Z}$-sublattice $L_{1}$ of $L$, vectors $w_{1}, w_{2}$ of $L_{1}$, and vectors $v_{1}, v_{2}$ of $L$. Suppose $w_{1}=v_{1}$ and $w_{2}=v_{2}$. Then $\left\langle w_{1}, w_{2}\right\rangle=\left\langle v_{1}, v_{2}\right\rangle$.

Let $L$ be an integral $\mathbb{Z}$-lattice. Note that every $\mathbb{Z}$-sublattice of $L$ is integral. Let $L$ be a positive definite $\mathbb{Z}$-lattice. Let us observe that every $\mathbb{Z}$-sublattice of $L$ is positive definite.

Let $V, W$ be vector space structures over $\mathbb{Z}^{\mathrm{R}}$.
An $\mathbb{R}$-form of $V$ and $W$ is a function from (the carrier of $V$ ) $\times$ (the carrier of $W$ ) into the carrier of $\mathbb{R}_{F}$. The functor $\operatorname{NulFrForm}(V, W)$ yielding an $\mathbb{R}$-form of $V$ and $W$ is defined by the term
(Def. 10) (the carrier of $V) \times($ the carrier of $W) \longmapsto 0_{\mathbb{R}_{F}}$.
Let $V, W$ be non empty vector space structures over $\mathbb{Z}^{\mathrm{R}}$ and $f, g$ be $\mathbb{R}$-forms of $V$ and $W$. The functor $f+g$ yielding an $\mathbb{R}$-form of $V$ and $W$ is defined by
(Def. 11) for every vector $v$ of $V$ and for every vector $w$ of $W, i t(v, w)=f(v, w)+$ $g(v, w)$.
Let $f$ be an $\mathbb{R}$-form of $V$ and $W$ and $a$ be an element of $\mathbb{R}_{F}$. The functor $a \cdot f$ yielding an $\mathbb{R}$-form of $V$ and $W$ is defined by
(Def. 12) for every vector $v$ of $V$ and for every vector $w$ of $W, i t(v, w)=a \cdot f(v, w)$. The functor $-f$ yielding an $\mathbb{R}$-form of $V$ and $W$ is defined by
(Def. 13) for every vector $v$ of $V$ and for every vector $w$ of $W$, it $(v, w)=-f(v, w)$.
One can verify that the functor $-f$ is defined by the term
(Def. 14) $\quad\left(-1_{\mathbb{R}_{F}}\right) \cdot f$.
Let $f, g$ be $\mathbb{R}$-forms of $V$ and $W$. The functor $f-g$ yielding an $\mathbb{R}$-form of $V$ and $W$ is defined by the term
(Def. 15) $f+-g$.
Observe that the functor $f-g$ is defined by
(Def. 16) for every vector $v$ of $V$ and for every vector $w$ of $W, i t(v, w)=f(v, w)-$ $g(v, w)$.
Let us note that the functor $f+g$ is commutative.
Now we state the propositions:
(39) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, and an $\mathbb{R}$-form $f$ of $V$ and $W$. Then $f+\operatorname{NulFrForm}(V, W)=f$.
(40) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, and $\mathbb{R}$-forms $f, g, h$ of $V$ and $W$. Then $(f+g)+h=f+(g+h)$.
(41) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, and an $\mathbb{R}$-form $f$ of $V$ and $W$. Then $f-f=\operatorname{NulFrForm}(V, W)$.
(42) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, an element $a$ of $\mathbb{R}_{\mathrm{F}}$, and $\mathbb{R}$-forms $f, g$ of $V$ and $W$. Then $a \cdot(f+g)=a \cdot f+a \cdot g$.
Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, elements $a, b$ of $\mathbb{R}_{F}$, and an $\mathbb{R}$-form $f$ of $V$ and $W$. Now we state the propositions:

$$
\begin{align*}
& \text { (43) }(a+b) \cdot f=a \cdot f+b \cdot f \text {. }  \tag{43}\\
& \text { (44) }(a \cdot b) \cdot f=a \cdot(b \cdot f) \text {. } \\
& \text { (45) Let us consider non empty vector space structures } V, W \text { over } \mathbb{Z}^{\mathrm{R}} \text {, and } \\
& \text { an } \mathbb{R} \text {-form } f \text { of } V \text { and } W \text {. Then } 1_{\mathbb{R}_{F}} \cdot f=f \text {. }
\end{align*}
$$

Let $V$ be a vector space structure over $\mathbb{Z}^{\mathrm{R}}$.
An $\mathbb{R}$-functional of $V$ is a function from the carrier of $V$ into the carrier of $\mathbb{R}_{\mathrm{F}}$. Let $V$ be a non empty vector space structure over $\mathbb{Z}^{\mathrm{R}}$ and $f, g$ be $\mathbb{R}$ functionals of $V$. The functor $f+g$ yielding an $\mathbb{R}$-functional of $V$ is defined by
(Def. 17) for every element $x$ of $V, i t(x)=f(x)+g(x)$.
Let $f$ be an $\mathbb{R}$-functional of $V$. The functor $-f$ yielding an $\mathbb{R}$-functional of $V$ is defined by
(Def. 18) for every element $x$ of $V$, it $(x)=-f(x)$.
Let $f, g$ be $\mathbb{R}$-functionals of $V$. The functor $f-g$ yielding an $\mathbb{R}$-functional of $V$ is defined by the term
(Def. 19) $f+-g$.
Let $v$ be an element of $\mathbb{R}_{\mathrm{F}}$ and $f$ be an $\mathbb{R}$-functional of $V$. The functor $v \cdot f$ yielding an $\mathbb{R}$-functional of $V$ is defined by
(Def. 20) for every element $x$ of $V, i t(x)=v \cdot f(x)$.
Let $V$ be a vector space structure over $\mathbb{Z}^{\mathrm{R}}$. The functor $0 \operatorname{FrFunctional}(V)$ yielding an $\mathbb{R}$-functional of $V$ is defined by the term
(Def. 21) $\quad \Omega_{V} \longmapsto 0_{\mathbb{R}_{\mathrm{F}}}$.
Let $V$ be a non empty vector space structure over $\mathbb{Z}^{\mathrm{R}}$ and $F$ be an $\mathbb{R}$ functional of $V$. We say that $F$ is homogeneous if and only if
(Def. 22) for every vector $x$ of $V$ and for every scalar $r$ of $V, F(r \cdot x)=r \cdot F(x)$.
We say that $F$ is 0 -preserving if and only if
(Def. 23) $\quad F\left(0_{V}\right)=0_{\mathbb{R}_{\boldsymbol{F}}}$.
Let $V$ be a $\mathbb{Z}$-module. Note that every $\mathbb{R}$-functional of $V$ which is homogeneous is also 0 -preserving.

Let $V$ be a non empty vector space structure over $\mathbb{Z}^{\mathrm{R}}$. One can verify that $0 \operatorname{FrFunctional}(V)$ is additive and $0 \operatorname{FrFunctional}(V)$ is homogeneous and $0 \mathrm{FrFunctional}(V)$ is 0 -preserving and there exists an $\mathbb{R}$-functional of $V$ which is additive, homogeneous, and 0 -preserving.

Now we state the propositions:
(46) Let us consider a non empty vector space structure $V$ over $\mathbb{Z}^{\mathrm{R}}$, and $\mathbb{R}$-functionals $f, g$ of $V$. Then $f+g=g+f$.
(47) Let us consider a non empty vector space structure $V$ over $\mathbb{Z}^{\mathrm{R}}$, and $\mathbb{R}$-functionals $f, g, h$ of $V$. Then $(f+g)+h=f+(g+h)$.
(48) Let us consider a non empty vector space structure $V$ over $\mathbb{Z}^{\mathrm{R}}$, and an element $x$ of $V$. Then $(0 \operatorname{FrFunctional}(V))(x)=0_{\mathbb{R}_{F}}$.
Let us consider a non empty vector space structure $V$ over $\mathbb{Z}^{\mathrm{R}}$ and an $\mathbb{R}$ functional $f$ of $V$. Now we state the propositions:
(49) $\quad f+0 \operatorname{FrFunctional}(V)=f$.
(50) $\quad f-f=0 \operatorname{FrFunctional}(V)$.
(51) Let us consider a non empty vector space structure $V$ over $\mathbb{Z}^{\mathrm{R}}$, an element $r$ of $\mathbb{R}_{\mathrm{F}}$, and $\mathbb{R}$-functionals $f, g$ of $V$. Then $r \cdot(f+g)=r \cdot f+r \cdot g$.
Let us consider a non empty vector space structure $V$ over $\mathbb{Z}^{\mathrm{R}}$, elements $r$, $s$ of $\mathbb{R}_{\mathrm{F}}$, and an $\mathbb{R}$-functional $f$ of $V$. Now we state the propositions:
(52) $\quad(r+s) \cdot f=r \cdot f+s \cdot f$.
(53) $(r \cdot s) \cdot f=r \cdot(s \cdot f)$.
(54) Let us consider a non empty vector space structure $V$ over $\mathbb{Z}^{\mathrm{R}}$, and an $\mathbb{R}$-functional $f$ of $V$. Then $1_{\mathbb{R}_{F}} \cdot f=f$.
Let $V$ be a non empty vector space structure over $\mathbb{Z}^{\mathrm{R}}$ and $f, g$ be additive $\mathbb{R}$-functionals of $V$. Observe that $f+g$ is additive.

Let $f$ be an additive $\mathbb{R}$-functional of $V$. One can check that $-f$ is additive.
Let $v$ be an element of $\mathbb{R}_{\mathrm{F}}$. Let us note that $v \cdot f$ is additive.
Let $f, g$ be homogeneous $\mathbb{R}$-functionals of $V$. Let us observe that $f+g$ is homogeneous.

Let $f$ be a homogeneous $\mathbb{R}$-functional of $V$. Note that $-f$ is homogeneous.
Let $v$ be an element of $\mathbb{R}_{F}$. Observe that $v \cdot f$ is homogeneous.
Let $V, W$ be non empty vector space structures over $\mathbb{Z}^{\mathrm{R}}, f$ be an $\mathbb{R}$-form of $V$ and $W$, and $v$ be a vector of $V$. The functor $\operatorname{FrFunctionalFAF}(f, v)$ yielding an $\mathbb{R}$-functional of $W$ is defined by the term
(Def. 24) (curry $f$ ) $(v)$.
Let $w$ be a vector of $W$. The functor $\operatorname{FrFunctionalSAF}(f, w)$ yielding an $\mathbb{R}$ functional of $V$ is defined by the term
(Def. 25) (curry' $f)(w)$.
Now we state the propositions:
(55) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, an $\mathbb{R}$ form $f$ of $V$ and $W$, and a vector $v$ of $V$. Then
(i) dom $\operatorname{FrFunctionalFAF}(f, v)=$ the carrier of $W$, and
(ii) rng $\operatorname{FrFunctionalFAF}(f, v) \subseteq$ the carrier of $\mathbb{R}_{F}$, and
(iii) for every vector $w$ of $W$, $(\operatorname{FrFunctionalFAF}(f, v))(w)=f(v, w)$.
(56) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, an $\mathbb{R}$ form $f$ of $V$ and $W$, and a vector $w$ of $W$. Then
(i) dom $\operatorname{FrFunctionalSAF}(f, w)=$ the carrier of $V$, and
(ii) rng $\operatorname{FrFunctionalSAF}(f, w) \subseteq$ the carrier of $\mathbb{R}_{\mathrm{F}}$, and
(iii) for every vector $v$ of $V$, $(\operatorname{FrFunctionalSAF}(f, w))(v)=f(v, w)$.
(57) Let us consider a non empty vector space structure $V$ over $\mathbb{Z}^{\mathrm{R}}$, and an element $x$ of $V$. Then $(0 \operatorname{FrFunctional}(V))(x)=0_{\mathbb{R}_{F}}$.
(58) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, and a vector $v$ of $V$. Then $\operatorname{FrFunctionalFAF}(\operatorname{NulFrForm}(V, W), v)=$ $0 \mathrm{FrFunctional}(W)$. The theorem is a consequence of (55).
(59) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, and a vector $w$ of $W$. Then FrFunctionalSAF(NulFrForm $(V, W), w)=$ 0 FrFunctional $(V)$. The theorem is a consequence of (56).
(60) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}, \mathbb{R}$ forms $f, g$ of $V$ and $W$, and a vector $w$ of $W$. Then $\operatorname{FrFunctionalSAF~}(f+$ $g, w)=\operatorname{FrFunctionalSAF}(f, w)+\operatorname{FrFunctionalSAF}(g, w)$. The theorem is a consequence of (56).
(61) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}, \mathbb{R}$ forms $f, g$ of $V$ and $W$, and a vector $v$ of $V$. Then $\operatorname{FrFunctionalFAF~}(f+$ $g, v)=\operatorname{FrFunctionalFAF}(f, v)+\operatorname{FrFunctionalFAF}(g, v)$. The theorem is a consequence of (55).
(62) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, an $\mathbb{R}$ form $f$ of $V$ and $W$, an element $a$ of $\mathbb{R}_{\mathrm{F}}$, and a vector $w$ of $W$. Then $\operatorname{FrFunctionalSAF}(a \cdot f, w)=a \cdot \operatorname{FrFunctionalSAF}(f, w)$. The theorem is a consequence of (56).
(63) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, an $\mathbb{R}$ form $f$ of $V$ and $W$, an element $a$ of $\mathbb{R}_{F}$, and a vector $v$ of $V$. Then $\operatorname{FrFunctionalFAF}(a \cdot f, v)=a \cdot \operatorname{FrFunctionalFAF}(f, v)$. The theorem is a consequence of (55).
(64) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, an $\mathbb{R}$ form $f$ of $V$ and $W$, and a vector $w$ of $W$. Then $\operatorname{FrFunctionalSAF~}(-f, w)=$ -FrFunctionalSAF $(f, w)$. The theorem is a consequence of (56).
(65) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, an $\mathbb{R}$ form $f$ of $V$ and $W$, and a vector $v$ of $V$. Then $\operatorname{FrFunctionalFAF~}(-f, v)=$ $-\operatorname{FrFunctionalFAF}(f, v)$. The theorem is a consequence of (55).
(66) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}, \mathbb{R}$ forms $f, g$ of $V$ and $W$, and a vector $w$ of $W$. Then $\operatorname{FrFunctionalSAF~}(f-$
$g, w)=\operatorname{FrFunctionalSAF}(f, w)-\operatorname{FrFunctionalSAF}(g, w)$. The theorem is a consequence of (56).
(67) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}, \mathbb{R}$ forms $f, g$ of $V$ and $W$, and a vector $v$ of $V$. Then $\operatorname{FrFunctionalFAF}(f-$ $g, v)=\operatorname{FrFunctionalFAF}(f, v)-\operatorname{FrFunctionalFAF}(g, v)$. The theorem is a consequence of (55).
Let $V, W$ be non empty vector space structures over $\mathbb{Z}^{\mathrm{R}}, f$ be an $\mathbb{R}$-functional of $V$, and $g$ be an $\mathbb{R}$-functional of $W$. The functor $\operatorname{FrFormFunctional}(f, g)$ yielding an $\mathbb{R}$-form of $V$ and $W$ is defined by
(Def. 26) for every vector $v$ of $V$ and for every vector $w$ of $W, i t(v, w)=f(v) \cdot g(w)$.
(68) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, an $\mathbb{R}$ functional $f$ of $V$, a vector $v$ of $V$, and a vector $w$ of $W$.
Then $(\operatorname{FrFormFunctional}(f, 0 \operatorname{FrFunctional}(W)))(v, w)=0_{\mathbb{Z}^{\mathrm{R}}}$.
(69) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, an $\mathbb{R}$ functional $g$ of $W$, a vector $v$ of $V$, and a vector $w$ of $W$. Then $(\operatorname{FrFormFunctional}(0 \operatorname{FrFunctional}(V), g))(v, w)=0_{\mathbb{Z}^{\mathrm{R}}}$.
(70) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, and an $\mathbb{R}$-functional $f$ of $V$. Then $\operatorname{FrFormFunctional}(f, 0 \operatorname{FrFunctional}(W))=$ $\operatorname{NulFrForm}(V, W)$. The theorem is a consequence of (68).
(71) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, and an $\mathbb{R}$-functional $g$ of $W$. Then FrFormFunctional $(0 \operatorname{FrFunctional}(V), g)=$ $\operatorname{NulFrForm}(V, W)$. The theorem is a consequence of (69).
(72) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, an $\mathbb{R}$ functional $f$ of $V$, an $\mathbb{R}$-functional $g$ of $W$, and a vector $v$ of $V$. Then $\operatorname{FrFunctionalFAF}(\operatorname{FrFormFunctional}(f, g), v)=f(v) \cdot g$. The theorem is a consequence of ( 55 ).
(73) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, an $\mathbb{R}$ functional $f$ of $V$, an $\mathbb{R}$-functional $g$ of $W$, and a vector $w$ of $W$. Then FrFunctionalSAF $(\operatorname{FrFormFunctional}(f, g), w)=g(w) \cdot f$. The theorem is a consequence of (56).

## 2. Bilinear Forms over Field of Reals and Their Properties

Let $V, W$ be non empty vector space structures over $\mathbb{Z}^{\mathrm{R}}$ and $f$ be an $\mathbb{R}$-form of $V$ and $W$. We say that $f$ is additive w.r.t. second argument if and only if
(Def. 27) for every vector $v$ of $V$, $\operatorname{FrFunctionalFAF}(f, v)$ is additive.
We say that $f$ is additive w.r.t. first argument if and only if
(Def. 28) for every vector $w$ of $W$, $\operatorname{FrFunctionalSAF}(f, w)$ is additive.

We say that $f$ is homogeneous w.r.t. second argument if and only if
(Def. 29) for every vector $v$ of $V, \operatorname{FrFunctionalFAF}(f, v)$ is homogeneous.
We say that $f$ is homogeneous w.r.t. first argument if and only if
(Def. 30) for every vector $w$ of $W, \operatorname{FrFunctionalSAF}(f, w)$ is homogeneous.
Observe that $\operatorname{NulFrForm}(V, W)$ is additive w.r.t. second argument and
$\operatorname{NulFrForm}(V, W)$ is additive w.r.t. first argument and there exists an $\mathbb{R}$ form of $V$ and $W$ which is additive w.r.t. second argument and additive w.r.t. first argument and $\operatorname{NulFrForm}(V, W)$ is homogeneous w.r.t. second argument and $\operatorname{NulFrForm}(V, W)$ is homogeneous w.r.t. first argument.

There exists an $\mathbb{R}$-form of $V$ and $W$ which is additive w.r.t. second argument, homogeneous w.r.t. second argument, additive w.r.t. first argument, and homogeneous w.r.t. first argument.

An $\mathbb{R}$-bilinear form of $V$ and $W$ is an additive w.r.t. first argument, homogeneous w.r.t. first argument, additive w.r.t. second argument, homogeneous w.r.t. second argument $\mathbb{R}$-form of $V$ and $W$. Let $f$ be an additive w.r.t. second argument $\mathbb{R}$-form of $V$ and $W$ and $v$ be a vector of $V$. One can check that FrFunctionalFAF $(f, v)$ is additive.

Let $f$ be an additive w.r.t. first argument $\mathbb{R}$-form of $V$ and $W$ and $w$ be a vector of $W$. Observe that FrFunctionalSAF $(f, w)$ is additive.

Let $f$ be a homogeneous w.r.t. second argument $\mathbb{R}$-form of $V$ and $W$ and $v$ be a vector of $V$. One can check that $\operatorname{FrFunctionalFAF}(f, v)$ is homogeneous.

Let $f$ be a homogeneous w.r.t. first argument $\mathbb{R}$-form of $V$ and $W$ and $w$ be a vector of $W$. Observe that FrFunctionalSAF $(f, w)$ is homogeneous.

Let $f$ be an $\mathbb{R}$-functional of $V$ and $g$ be an additive $\mathbb{R}$-functional of $W$. Observe that $\operatorname{FrFormFunctional}(f, g)$ is additive w.r.t. second argument.

Let $f$ be an additive $\mathbb{R}$-functional of $V$ and $g$ be an $\mathbb{R}$-functional of $W$. One can check that $\operatorname{FrFormFunctional}(f, g)$ is additive w.r.t. first argument.

Let $f$ be an $\mathbb{R}$-functional of $V$ and $g$ be a homogeneous $\mathbb{R}$-functional of $W$. Observe that $\operatorname{FrFormFunctional}(f, g)$ is homogeneous w.r.t. second argument.

Let $f$ be a homogeneous $\mathbb{R}$-functional of $V$ and $g$ be an $\mathbb{R}$-functional of $W$. One can check that $\operatorname{FrFormFunctional}(f, g)$ is homogeneous w.r.t. first argument.

Let $V$ be a non trivial vector space structure over $\mathbb{Z}^{\mathrm{R}}, W$ be a non empty vector space structure over $\mathbb{Z}^{\mathrm{R}}$, and $f$ be an $\mathbb{R}$-functional of $V$. One can verify that $\operatorname{FrFormFunctional}(f, g)$ is non trivial and $\operatorname{FrFormFunctional}(f, g)$ is non trivial.

Let $F$ be an $\mathbb{R}$-functional of $V$. We say that $F$ is 0 -preserving if and only if (Def. 31) $\quad F\left(0_{V}\right)=0_{\mathbb{R}_{\mathrm{F}}}$.

Let $V$ be a $\mathbb{Z}$-module. One can check that every $\mathbb{R}$-functional of $V$ which is homogeneous is also 0 -preserving.

Let $V$ be a non empty vector space structure over $\mathbb{Z}^{\mathrm{R}}$. Let us observe that 0 FrFunctional $(V)$ is 0 -preserving and there exists an $\mathbb{R}$-functional of $V$ which is additive, homogeneous, and 0-preserving.

Let $V$ be a non trivial, free $\mathbb{Z}$-module. Note that there exists an $\mathbb{R}$-functional of $V$ which is additive, homogeneous, non constant, and non trivial.
(74) Let us consider a non trivial, free $\mathbb{Z}$-module $V$, and a non constant, 0preserving $\mathbb{R}$-functional $f$ of $V$. Then there exists a vector $v$ of $V$ such that
(i) $v \neq 0_{V}$, and
(ii) $f(v) \neq 0_{\mathbb{R}_{F}}$.

Let $V, W$ be non trivial, free $\mathbb{Z}$-modules, $f$ be a non constant, 0 -preserving $\mathbb{R}$-functional of $V$, and $g$ be a non constant, 0 -preserving $\mathbb{R}$-functional of $W$. Note that FrFormFunctional $(f, g)$ is non constant.

Let $V$ be a non empty vector space structure over $\mathbb{Z}^{\mathrm{R}}$.
An $\mathbb{R}$-linear functional of $V$ is an additive, homogeneous $\mathbb{R}$-functional of $V$. Let $V, W$ be non trivial, free $\mathbb{Z}$-modules. Observe that there exists an $\mathbb{R}$ form of $V$ and $W$ which is non trivial, non constant, additive w.r.t. second argument, homogeneous w.r.t. second argument, additive w.r.t. first argument, and homogeneous w.r.t. first argument.

Let $V, W$ be non empty vector space structures over $\mathbb{Z}^{\mathrm{R}}$ and $f, g$ be additive w.r.t. first argument $\mathbb{R}$-forms of $V$ and $W$. Let us observe that $f+g$ is additive w.r.t. first argument. Let $f, g$ be additive w.r.t. second argument $\mathbb{R}$-forms of $V$ and $W$. One can check that $f+g$ is additive w.r.t. second argument.

Let $f$ be an additive w.r.t. first argument $\mathbb{R}$-form of $V$ and $W$ and $a$ be an element of $\mathbb{R}_{\mathrm{F}}$. Let us observe that $a \cdot f$ is additive w.r.t. first argument.

Let $f$ be an additive w.r.t. second argument $\mathbb{R}$-form of $V$ and $W$. Note that $a \cdot f$ is additive w.r.t. second argument.

Let $f$ be an additive w.r.t. first argument $\mathbb{R}$-form of $V$ and $W$. Let us observe that $-f$ is additive w.r.t. first argument.

Let $f$ be an additive w.r.t. second argument $\mathbb{R}$-form of $V$ and $W$. Let us observe that $-f$ is additive w.r.t. second argument.

Let $f, g$ be additive w.r.t. first argument $\mathbb{R}$-forms of $V$ and $W$. Observe that $f-g$ is additive w.r.t. first argument.

Let $f, g$ be additive w.r.t. second argument $\mathbb{R}$-forms of $V$ and $W$. One can check that $f-g$ is additive w.r.t. second argument.

Let $f, g$ be homogeneous w.r.t. first argument $\mathbb{R}$-forms of $V$ and $W$. Observe that $f+g$ is homogeneous w.r.t. first argument.

Let $f, g$ be homogeneous w.r.t. second argument $\mathbb{R}$-forms of $V$ and $W$. One can verify that $f+g$ is homogeneous w.r.t. second argument.

Let $f$ be a homogeneous w.r.t. first argument $\mathbb{R}$-form of $V$ and $W$ and $a$ be an element of $\mathbb{R}_{\mathrm{F}}$. Observe that $a \cdot f$ is homogeneous w.r.t. first argument.

Let $f$ be a homogeneous w.r.t. second argument $\mathbb{R}$-form of $V$ and $W$. One can check that $a \cdot f$ is homogeneous w.r.t. second argument.

Let $f$ be a homogeneous w.r.t. first argument $\mathbb{R}$-form of $V$ and $W$. Observe that $-f$ is homogeneous w.r.t. first argument. Let $f$ be a homogeneous w.r.t. second argument $\mathbb{R}$-form of $V$ and $W$. Observe that $-f$ is homogeneous w.r.t. second argument.

Let $f, g$ be homogeneous w.r.t. first argument $\mathbb{R}$-forms of $V$ and $W$. Let us note that $f-g$ is homogeneous w.r.t. first argument.

Let $f, g$ be homogeneous w.r.t. second argument $\mathbb{R}$-forms of $V$ and $W$. One can verify that $f-g$ is homogeneous w.r.t. second argument.
(75) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, vectors $v, u$ of $V$, a vector $w$ of $W$, and an $\mathbb{R}$-form $f$ of $V$ and $W$. If $f$ is additive w.r.t. first argument, then $f(v+u, w)=f(v, w)+f(u, w)$. The theorem is a consequence of (56).
(76) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, a vector $v$ of $V$, vectors $u, w$ of $W$, and an $\mathbb{R}$-form $f$ of $V$ and $W$. If $f$ is additive w.r.t. second argument, then $f(v, u+w)=f(v, u)+f(v, w)$. The theorem is a consequence of (55).
(77) Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, vectors $v, u$ of $V$, vectors $w, t$ of $W$, and an additive w.r.t. first argument, additive w.r.t. second argument $\mathbb{R}$-form $f$ of $V$ and $W$. Then $f(v+u, w+t)=$ $f(v, w)+f(v, t)+(f(u, w)+f(u, t))$. The theorem is a consequence of (75) and (76).
(78) Let us consider right zeroed, non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, an additive w.r.t. second argument $\mathbb{R}$-form $f$ of $V$ and $W$, and a vector $v$ of $V$. Then $f\left(v, 0_{W}\right)=0_{\mathbb{Z}^{R}}$. The theorem is a consequence of (76).
(79) Let us consider right zeroed, non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, an additive w.r.t. first argument $\mathbb{R}$-form $f$ of $V$ and $W$, and a vector $w$ of $W$. Then $f\left(0_{V}, w\right)=0_{\mathbb{Z}^{\mathrm{R}}}$. The theorem is a consequence of (75).

Let us consider non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, a vector $v$ of $V$, a vector $w$ of $W$, an element $a$ of $\mathbb{Z}^{\mathrm{R}}$, and an $\mathbb{R}$-form $f$ of $V$ and $W$. Now we state the propositions:
(80) If $f$ is homogeneous w.r.t. first argument, then $f(a \cdot v, w)=a \cdot f(v, w)$.

The theorem is a consequence of (56).
(81) If $f$ is homogeneous w.r.t. second argument, then $f(v, a \cdot w)=a \cdot f(v, w)$. The theorem is a consequence of (55).
(82) Let us consider add-associative, right zeroed, right complementable, vector distributive, scalar distributive, scalar associative, scalar unital, non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, a homogeneous w.r.t. first argument $\mathbb{R}$-form $f$ of $V$ and $W$, and a vector $w$ of $W$. Then $f\left(0_{V}, w\right)=0_{\mathbb{R}_{F}}$. The theorem is a consequence of (80).
(83) Let us consider add-associative, right zeroed, right complementable, vector distributive, scalar distributive, scalar associative, scalar unital, non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, a homogeneous w.r.t. second argument $\mathbb{R}$-form $f$ of $V$ and $W$, and a vector $v$ of $V$. Then $f\left(v, 0_{W}\right)=0_{\mathbb{R}_{F}}$. The theorem is a consequence of (81).
(84) Let us consider $\mathbb{Z}$-modules $V, W$, vectors $v, u$ of $V$, a vector $w$ of $W$, and an additive w.r.t. first argument, homogeneous w.r.t. first argument $\mathbb{R}$-form $f$ of $V$ and $W$. Then $f(v-u, w)=f(v, w)-f(u, w)$. The theorem is a consequence of (75) and (80).
(85) Let us consider $\mathbb{Z}$-modules $V, W$, a vector $v$ of $V$, vectors $w, t$ of $W$, and an additive w.r.t. second argument, homogeneous w.r.t. second argument $\mathbb{R}$-form $f$ of $V$ and $W$. Then $f(v, w-t)=f(v, w)-f(v, t)$. The theorem is a consequence of $(76)$ and (81).
(86) Let us consider $\mathbb{Z}$-modules $V, W$, vectors $v, u$ of $V$, vectors $w, t$ of $W$, and an $\mathbb{R}$-bilinear form $f$ of $V$ and $W$. Then $f(v-u, w-t)=f(v, w)-$ $f(v, t)-(f(u, w)-f(u, t))$. The theorem is a consequence of (84) and (85).
(87) Let us consider add-associative, right zeroed, right complementable, vector distributive, scalar distributive, scalar associative, scalar unital, non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, vectors $v, u$ of $V$, vectors $w, t$ of $W$, elements $a, b$ of $\mathbb{Z}^{\mathrm{R}}$, and an $\mathbb{R}$-bilinear form $f$ of $V$ and $W$. Then $f(v+a \cdot u, w+b \cdot t)=f(v, w)+b \cdot f(v, t)+(a \cdot f(u, w)+a \cdot(b \cdot f(u, t)))$. The theorem is a consequence of (77), (81), and (80).
(88) Let us consider $\mathbb{Z}$-modules $V, W$, vectors $v, u$ of $V$, vectors $w, t$ of $W$, elements $a, b$ of $\mathbb{Z}^{\mathrm{R}}$, and an $\mathbb{R}$-bilinear form $f$ of $V$ and $W$. Then $f(v-a \cdot u, w-b \cdot t)=f(v, w)-b \cdot f(v, t)-(a \cdot f(u, w)-a \cdot(b \cdot f(u, t)))$. The theorem is a consequence of (86), (81), and (80).
(89) Let us consider right zeroed, non empty vector space structures $V, W$ over $\mathbb{Z}^{\mathrm{R}}$, and an $\mathbb{R}$-form $f$ of $V$ and $W$. Suppose $f$ is additive w.r.t. second argument or additive w.r.t. first argument. Then $f$ is constant if and only if for every vector $v$ of $V$ and for every vector $w$ of $W, f(v, w)=0_{\mathbb{Z}^{R}}$. The theorem is a consequence of (78) and (79).

## 3. Matrices of Bilinear Form over Field of Real Numbers

Let $V_{1}, V_{2}$ be finite rank, free $\mathbb{Z}$-modules, $b_{1}$ be an ordered basis of $V_{1}, b_{2}$ be an ordered basis of $V_{2}$, and $f$ be an $\mathbb{R}$-bilinear form of $V_{1}$ and $V_{2}$. The functor $\operatorname{Bilinear}\left(f, b_{1}, b_{2}\right)$ yielding a matrix over $\mathbb{R}_{\mathrm{F}}$ of dimension len $b_{1} \times \operatorname{len} b_{2}$ is defined by
(Def. 32) for every natural numbers $i, j$ such that $i \in \operatorname{dom} b_{1}$ and $j \in \operatorname{dom} b_{2}$ holds $i t_{i, j}=f\left(b_{1 i}, b_{2 j}\right)$.
Now we state the propositions:
(90) Let us consider a finite rank, free $\mathbb{Z}$-module $V$, an $\mathbb{R}$-linear functional $F$ of $V$, a finite sequence $y$ of elements of $V$, a finite sequence $x$ of elements of $\mathbb{Z}^{\mathrm{R}}$, and finite sequences $X, Y$ of elements of $\mathbb{R}_{\mathrm{F}}$. Suppose $X=x$ and len $y=\operatorname{len} x$ and len $X=\operatorname{len} Y$ and for every natural number $k$ such that $k \in \operatorname{Seg}$ len $x$ holds $Y(k)=F\left(y_{k}\right)$. Then $X \cdot Y=F\left(\sum \operatorname{lmlt}(x, y)\right)$.
Proof: Define $\mathcal{P}$ [finite sequence of elements of $V] \equiv$ for every finite sequence $x$ of elements of $\mathbb{Z}^{\mathrm{R}}$ for every finite sequences $X, Y$ of elements of $\mathbb{R}_{F}$ such that $X=x$ and len $\$_{1}=\operatorname{len} x$ and len $X=\operatorname{len} Y$ and for every natural number $k$ such that $k \in \operatorname{Seg} \operatorname{len} x$ holds $Y(k)=F\left(\$_{1 k}\right)$ holds $X \cdot Y=F\left(\sum \operatorname{lmlt}\left(x, \$_{1}\right)\right)$. For every finite sequence $y$ of elements of $V$ and for every element $w$ of $V$ such that $\mathcal{P}[y]$ holds $\mathcal{P}\left[y^{\sim}\langle w\rangle\right.$ ] by [4, (22), (39), (59)], [3, (11)]. $\mathcal{P}\left[\varepsilon_{\alpha}\right]$, where $\alpha$ is the carrier of $V$ by [17, (43)]. For every finite sequence $p$ of elements of $V, \mathcal{P}[p]$ from [6, Sch. 2].
(91) Let us consider finite rank, free $\mathbb{Z}$-modules $V_{1}, V_{2}$, an ordered basis $b_{2}$ of $V_{2}$, an ordered basis $b_{3}$ of $V_{2}$, an $\mathbb{R}$-bilinear form $f$ of $V_{1}$ and $V_{2}$, a vector $v_{1}$ of $V_{1}$, a vector $v_{2}$ of $V_{2}$, and finite sequences $X, Y$ of elements of $\mathbb{R}_{F}$. Suppose len $X=\operatorname{len} b_{2}$ and len $Y=\operatorname{len} b_{2}$ and for every natural number $k$ such that $k \in \operatorname{Seg}$ len $b_{2}$ holds $Y(k)=f\left(v_{1}, b_{2 k}\right)$ and $X=v_{2} \rightarrow b_{2}$. Then $Y \cdot X=f\left(v_{1}, v_{2}\right)$. The theorem is a consequence of (55) and (90).
(92) Let us consider finite rank, free $\mathbb{Z}$-modules $V_{1}, V_{2}$, an ordered basis $b_{1}$ of $V_{1}$, an $\mathbb{R}$-bilinear form $f$ of $V_{1}$ and $V_{2}$, a vector $v_{1}$ of $V_{1}$, a vector $v_{2}$ of $V_{2}$, and finite sequences $X, Y$ of elements of $\mathbb{R}_{F}$. Suppose len $X=\operatorname{len} b_{1}$ and len $Y=\operatorname{len} b_{1}$ and for every natural number $k$ such that $k \in \operatorname{Seg}$ len $b_{1}$ holds $Y(k)=f\left(b_{1 k}, v_{2}\right)$ and $X=v_{1} \rightarrow b_{1}$. Then $X \cdot Y=f\left(v_{1}, v_{2}\right)$. The theorem is a consequence of (56) and (90).
(93) Every matrix over $\mathbb{Z}^{R}$ is a matrix over $\mathbb{R}_{F}$.

Let $M$ be a matrix over $\mathbb{Z}^{\mathrm{R}}$. The functor $\mathbb{Z} 2 \mathbb{R}(M)$ yielding a matrix over $\mathbb{R}_{\mathrm{F}}$ is defined by the term
(Def. 33) $M$.

Let $n, m$ be natural numbers and $M$ be a matrix over $\mathbb{Z}^{\mathrm{R}}$ of dimension $n \times m$. Note that the functor $\mathbb{Z} 2 \mathbb{R}(M)$ yields a matrix over $\mathbb{R}_{F}$ of dimension $n \times m$. Let $n$ be a natural number and $M$ be a square matrix over $\mathbb{Z}^{\mathrm{R}}$ of dimension $n$. Let us note that the functor $\mathbb{Z} 2 \mathbb{R}(M)$ yields a square matrix over $\mathbb{R}_{F}$ of dimension $n$. Now we state the propositions:
(94) Let us consider natural numbers $m, l, n$, a matrix $S$ over $\mathbb{Z}^{\mathrm{R}}$ of dimension $l \times m$, a matrix $T$ over $\mathbb{Z}^{\mathrm{R}}$ of dimension $m \times n$, a matrix $S_{1}$ over $\mathbb{R}_{\mathrm{F}}$ of dimension $l \times m$, and a matrix $T_{1}$ over $\mathbb{R}_{\mathrm{F}}$ of dimension $m \times n$. If $S=S_{1}$ and $T=T_{1}$ and $0<l$ and $0<m$, then $S \cdot T=S_{1} \cdot T_{1}$.
Proof: Reconsider $S_{3}=S \cdot T$ as a matrix over $\mathbb{Z}^{\mathrm{R}}$ of dimension $l \times n$. Reconsider $S_{2}=S_{1} \cdot T_{1}$ as a matrix over $\mathbb{R}_{\mathrm{F}}$ of dimension $l \times n$. For every natural numbers $i, j$ such that $\langle i, j\rangle \in$ the indices of $S_{3}$ holds $S_{3 i, j}=S_{2 i, j}$ by [8, (87)], [13, (2), (3), (37)].
(95) Let us consider a natural number $n$. Then $I_{\mathbb{Z}^{\mathrm{R}}}^{n \times n}=I_{\mathbb{R}_{\mathrm{F}}}^{n \times n}$.
(96) Let us consider finite rank, free $\mathbb{Z}$-modules $V_{1}, V_{2}$, an ordered basis $b_{1}$ of $V_{1}$, an ordered basis $b_{2}$ of $V_{2}$, an ordered basis $b_{3}$ of $V_{2}$, and an $\mathbb{R}$-bilinear form $f$ of $V_{1}$ and $V_{2}$. Suppose $0<\operatorname{rank} V_{1}$. Then $\operatorname{Bilinear}\left(f, b_{1}, b_{3}\right)=$ Bilinear $\left(f, b_{1}, b_{2}\right) \cdot\left(\mathbb{Z} 2 \mathbb{R}\left(\operatorname{AutMt}\left(\mathrm{id}_{V_{2}}, b_{3}, b_{2}\right)\right)\right)^{\mathrm{T}}$.
Proof: Set $n=\operatorname{len} b_{2}$. Reconsider $I_{2}=\operatorname{AutMt}\left(\mathrm{id}_{V_{2}}, b_{3}, b_{2}\right)$ as a square matrix over $\mathbb{Z}^{\mathrm{R}}$ of dimension $n$. Reconsider $M_{1}=\mathbb{Z} 2 \mathbb{R}\left(I_{2}{ }^{\mathrm{T}}\right)$ as a square matrix over $\mathbb{R}_{\mathrm{F}}$ of dimension $n$. Set $M_{2}=\operatorname{Bilinear}\left(f, b_{1}, b_{2}\right) \cdot M_{1}$. For every natural numbers $i, j$ such that $\langle i, j\rangle \in$ the indices of $\operatorname{Bilinear}\left(f, b_{1}, b_{3}\right)$ holds $\left(\operatorname{Bilinear}\left(f, b_{1}, b_{3}\right)\right)_{i, j}=M_{2 i, j}$ by [8, (87)], [13, (1)], (91).
(97) Let us consider finite rank, free $\mathbb{Z}$-modules $V_{1}, V_{2}$, an ordered basis $b_{1}$ of $V_{1}$, an ordered basis $b_{2}$ of $V_{2}$, an ordered basis $b_{3}$ of $V_{1}$, and an $\mathbb{R}$-bilinear form $f$ of $V_{1}$ and $V_{2}$. Suppose $0<\operatorname{rank} V_{1}$. Then $\operatorname{Bilinear}\left(f, b_{3}, b_{2}\right)=$ $\mathbb{Z} 2 \mathbb{R}\left(\operatorname{AutMt}\left(\mathrm{id}_{V_{1}}, b_{3}, b_{1}\right)\right) \cdot \operatorname{Bilinear}\left(f, b_{1}, b_{2}\right)$.
Proof: Set $n=\operatorname{len} b_{3}$. Reconsider $I_{2}=\operatorname{AutMt}\left(\mathrm{id}_{V_{1}}, b_{3}, b_{1}\right)$ as a square matrix over $\mathbb{Z}^{\mathrm{R}}$ of dimension $n$. Reconsider $M_{1}=\mathbb{Z} 2 \mathbb{R}\left(I_{2}\right)$ as a square matrix over $\mathbb{R}_{\mathrm{F}}$ of dimension $n$. Set $M_{2}=M_{1} \cdot \operatorname{Bilinear}\left(f, b_{1}, b_{2}\right)$. For every natural numbers $i, j$ such that $\langle i, j\rangle \in$ the indices of $\operatorname{Bilinear}\left(f, b_{3}, b_{2}\right)$ holds (Bilinear $\left.\left(f, b_{3}, b_{2}\right)\right)_{i, j}=M_{2 i, j}$ by [8, (87)], [4, (1)], [13, (1)], (92).
(98) Let us consider a finite rank, free $\mathbb{Z}$-module $V$, ordered bases $b_{1}, b_{2}$ of $V$, and an $\mathbb{R}$-bilinear form $f$ of $V$ and $V$. Suppose $0<\operatorname{rank} V$. Then Bilinear $\left(f, b_{2}, b_{2}\right)=\mathbb{Z} 2 \mathbb{R}\left(\operatorname{AutMt}\left(\mathrm{id}_{V}, b_{2}, b_{1}\right)\right) \cdot \operatorname{Bilinear}\left(f, b_{1}, b_{1}\right) \cdot(\mathbb{Z} 2 \mathbb{R}($ AutMt $\left.\left.\left(\operatorname{id}_{V}, b_{2}, b_{1}\right)\right)\right)^{\mathrm{T}}$. The theorem is a consequence of (97) and (96).
Let us consider a finite rank, free $\mathbb{Z}$-module $V$, ordered bases $b_{1}, b_{2}$ of $V$, and a square matrix $M$ over $\mathbb{R}_{F}$ of dimension rank $V$.

Let us assume that $M=\operatorname{AutMt}\left(\mathrm{id}_{V}, b_{1}, b_{2}\right)$. Now we state the propositions:
(99) (i) Det $M=1$ and $\operatorname{Det} M^{\mathrm{T}}=1$, or
(ii) $\operatorname{Det} M=-1$ and $\operatorname{Det} M^{\mathrm{T}}=-1$.

The theorem is a consequence of (94) and (95).
(100) $|\operatorname{Det} M|=1$. The theorem is a consequence of (99).

Let us consider a finite rank, free $\mathbb{Z}$-module $V$, ordered bases $b_{1}, b_{2}$ of $V$, and an $\mathbb{R}$-bilinear form $f$ of $V$ and $V$. Now we state the propositions:
(101) Det $\operatorname{Bilinear}\left(f, b_{2}, b_{2}\right)=\operatorname{Det} \operatorname{Bilinear}\left(f, b_{1}, b_{1}\right)$. The theorem is a consequence of (98) and (99).
(102) $\left|\operatorname{Det} \operatorname{Bilinear}\left(f, b_{2}, b_{2}\right)\right|=\left|\operatorname{Det} \operatorname{Bilinear}\left(f, b_{1}, b_{1}\right)\right|$.

Let $V$ be a finite rank, free $\mathbb{Z}$-module, $f$ be an $\mathbb{R}$-bilinear form of $V$ and $V$, and $b$ be an ordered basis of $V$. The functor $\operatorname{GramMatrix}(f, b)$ yielding a square matrix over $\mathbb{R}_{\mathrm{F}}$ of dimension rank $V$ is defined by the term
(Def. 34) Bilinear $(f, b, b)$.
The functor $\operatorname{GramDet}(f)$ yielding an element of $\mathbb{R}_{\mathrm{F}}$ is defined by
(Def. 35) for every ordered basis $b$ of $V$, it $=\operatorname{Det} \operatorname{GramMatrix}(f, b)$.
Let $L$ be a $\mathbb{Z}$-lattice. The functor InnerProduct $L$ yielding an $\mathbb{R}$-form of $L$ and $L$ is defined by the term
(Def. 36) the scalar product of $L$.
One can check that InnerProduct $L$ is additive w.r.t. first argument, homogeneous w.r.t. first argument, additive w.r.t. second argument, and homogeneous w.r.t. second argument.

Let $b$ be an ordered basis of $L$. The functor $\operatorname{GramMatrix}(b)$ yielding a square matrix over $\mathbb{R}_{\mathrm{F}}$ of dimension $\operatorname{dim}(L)$ is defined by the term
(Def. 37) GramMatrix(InnerProduct $L, b$ ).
The functor $\operatorname{GramDet}(L)$ yielding an element of $\mathbb{R}_{\mathrm{F}}$ is defined by the term
(Def. 38) GramDet(InnerProduct $L$ ).
(103) Let us consider an integral $\mathbb{Z}$-lattice $L$. Then InnerProduct $L$ is a bilinear form of $L, L$.
Proof: For every object $z$ such that $z \in($ the carrier of $L) \times($ the carrier of $L$ ) holds (InnerProduct $L)(z) \in$ the carrier of $\mathbb{Z}^{\mathrm{R}}$. Reconsider $f=$ InnerProduct $L$ as a form of $L, L$. For every vector $v$ of $L, f(\cdot, v)$ is additive by [2, (70)], (8). For every vector $v$ of $L, f(\cdot, v)$ is homogeneous by [2, (70)], (9). For every vector $v$ of $L, f(v, \cdot)$ is additive by [2, (69)], (8). For every vector $v$ of $L, f(v, \cdot)$ is homogeneous by [2, (69)], (9).
(104) Let us consider an integral $\mathbb{Z}$-lattice $L$, and an ordered basis $b$ of $L$. Then $\operatorname{GramMatrix}(b)$ is a square matrix over $\mathbb{Z}^{\mathrm{R}}$ of dimension $\operatorname{dim}(L)$.

Proof: For every natural numbers $i, j$ such that $\langle i, j\rangle \in$ the indices of GramMatrix $(b)$ holds $(\operatorname{GramMatrix}(b))_{i, j} \in$ the carrier of $\mathbb{Z}^{\mathrm{R}}$ by [8, (87)].

Let $L$ be an integral $\mathbb{Z}$-lattice. Note that $\operatorname{GramDet}(L)$ is integer.

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# Product Pre-Measure 

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#### Abstract

Summary. In this article we formalize in Mizar [5] product pre-measure on product sets of measurable sets. Although there are some approaches to construct product measure [22, [6], 9], [21, [25, we start it from $\sigma$-measure because existence of $\sigma$-measure on any semialgebras has been proved in 15. In this approach, we use some theorems for integrals.


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## 1. Preliminaries

Now we state the proposition:
(1) Let us consider non empty sets $A, A_{1}, A_{2}, B, B_{1}, B_{2}$. Then $A_{1} \times B_{1}$ misses $A_{2} \times B_{2}$ and $A \times B=A_{1} \times B_{1} \cup A_{2} \times B_{2}$ if and only if $A_{1}$ misses $A_{2}$ and $A=A_{1} \cup A_{2}$ and $B=B_{1}$ and $B=B_{2}$ or $B_{1}$ misses $B_{2}$ and $B=B_{1} \cup B_{2}$ and $A=A_{1}$ and $A=A_{2}$.
Let $C, D$ be non empty sets, $F$ be a sequence of $D^{C}$, and $n$ be a natural number. One can check that the functor $F(n)$ yields a function from $C$ into $D$.
(2) Let us consider sets $X, Y, A, B$, and objects $x, y$. Suppose $x \in X$ and $y \in Y$. Then $\chi_{A, X}(x) \cdot \chi_{B, Y}(y)=\chi_{A \times B, X \times Y}(x, y)$.
Let $A, B$ be sets. One can verify that $\chi_{A, B}$ is non-negative.
(3) Let us consider a non empty set $X$, a semialgebra $S$ of sets of $X$, a premeasure $P$ of $S$, an induced measure $m$ of $S$ and $P$, and an induced $\sigma$ measure $M$ of $S$ and $m$. Then $\operatorname{COM}(M)$ is complete on $\operatorname{COM}(\sigma$ (the field generated by $S), M)$.

The functor Intervals $\mathbb{R}_{\mathbb{R}}$ yielding a semialgebra of sets of $\mathbb{R}$ is defined by the term
(Def. 1) the set of all $I$ where $I$ is an interval.
Now we state the propositions:
(4) Halflines $\subseteq$ Intervals $_{\mathbb{R}}$.
(5) Let us consider a subset $I$ of $\mathbb{R}$. If $I$ is an interval, then $I \in$ the Borel sets.
(6) (i) $\sigma\left(\right.$ Intervals $\left._{\mathbb{R}}\right)=$ the Borel sets, and
(ii) $\sigma$ (the field generated by Intervals $\left.\mathbb{S}_{\mathbb{R}}\right)=$ the Borel sets.

The theorem is a consequence of (4) and (5).

## 2. Family of Semialgebras, Fields and Measures

Now we state the propositions:
(7) Let us consider sets $X_{1}, X_{2}$, a non empty family $S_{1}$ of subsets of $X_{1}$, and a non empty family $S_{2}$ of subsets of $X_{2}$. Then the set of all $a \times b$ where $a$ is an element of $S_{1}, b$ is an element of $S_{2}$ is a non empty family of subsets of $X_{1} \times X_{2}$.
(8) Let us consider sets $X, Y$, a family $M$ of subsets of $X$ with the empty element, and a family $N$ of subsets of $Y$ with the empty element. Then the set of all $A \times B$ where $A$ is an element of $M, B$ is an element of $N$ is a family of subsets of $X \times Y$ with the empty element. The theorem is a consequence of (7).
(9) Let us consider a set $X$, and disjoint valued finite sequences $O, T$ of elements of $X$. Suppose $\bigcup \operatorname{rng} O$ misses $\bigcup \operatorname{rng} T$. Then $O^{\wedge} T$ is a disjoint valued finite sequence of elements of $X$.
(10) Let us consider sets $X_{1}, X_{2}$, a semiring $S_{1}$ of $X_{1}$, and a semiring $S_{2}$ of $X_{2}$. Then the set of all $A \times B$ where $A$ is an element of $S_{1}, B$ is an element of $S_{2}$ is a semiring of $X_{1} \times X_{2}$.
(11) Let us consider sets $X_{1}, X_{2}$, a semialgebra $S_{1}$ of sets of $X_{1}$, and a semialgebra $S_{2}$ of sets of $X_{2}$. Then the set of all $A \times B$ where $A$ is an element of $S_{1}, B$ is an element of $S_{2}$ is a semialgebra of sets of $X_{1} \times X_{2}$. The theorem is a consequence of (10).
(12) Let us consider sets $X_{1}, X_{2}$, a field $O$ of subsets of $X_{1}$, and a field $T$ of subsets of $X_{2}$. Then the set of all $A \times B$ where $A$ is an element of $O$, $B$ is an element of $T$ is a semialgebra of sets of $X_{1} \times X_{2}$. The theorem is a consequence of (11).

Let $n$ be a non zero natural number and $X$ be a non-empty, $n$-element finite sequence.

A family of semialgebras of $X$ is an $n$-element finite sequence and is defined by
(Def. 2) for every natural number $i$ such that $i \in \operatorname{Seg} n$ holds $i t(i)$ is a semialgebra of sets of $X(i)$.
Let us observe that a family of semialgebras of $X$ is a $\cap$-closed yielding family of semirings of $X$. Now we state the proposition:
(13) Let us consider a non zero natural number $n$, a non-empty, $n$-element finite sequence $X$, a family $S$ of semialgebras of $X$, and a natural number $i$. If $i \in \operatorname{Seg} n$, then $X(i) \in S(i)$.
Let us consider a non-empty, 1-element finite sequence $X$ and a family $S$ of semialgebras of $X$. Now we state the propositions:
(14) the set of all $\Pi\langle s\rangle$ where $s$ is an element of $S(1)$ is a semialgebra of sets of the set of all $\langle x\rangle$ where $x$ is an element of $X(1)$. The theorem is a consequence of (13).
(15) SemiringProduct $(S)$ is a semialgebra of sets of $\Pi X$. The theorem is a consequence of (14).
(16) Let us consider sets $X_{1}, X_{2}$, a semialgebra $S_{1}$ of sets of $X_{1}$, and a semialgebra $S_{2}$ of sets of $X_{2}$. Then the set of all $s_{1} \times s_{2}$ where $s_{1}$ is an element of $S_{1}, s_{2}$ is an element of $S_{2}$ is a semialgebra of sets of $X_{1} \times X_{2}$.
(17) Let us consider a non zero natural number $n$, a non-empty, $n$-element finite sequence $X$, and a family $S$ of semialgebras of $X$. Then SemiringProduct $(S)$ is a semialgebra of sets of $\Pi X$.
Proof: Define $\mathcal{P}$ [non zero natural number] $\equiv$ for every non-empty, $\$_{1-}$ element finite sequence $X$ for every family $S$ of semialgebras of $X$, SemiringProduct $(S)$ is a semialgebra of sets of $\Pi X . \mathcal{P}[1]$. For every non zero natural number $k, \mathcal{P}[k]$ from [3, Sch. 10].
(18) Let us consider a non zero natural number $n$, a non-empty, $n$-element finite sequence $X_{8}$, a non-empty, 1-element finite sequence $X_{1}$, a family $S_{4}$ of semialgebras of $X_{8}$, and a family $S_{1}$ of semialgebras of $X_{1}$. Then SemiringProduct $\left(S_{4} \frown S_{1}\right)$ is a semialgebra of sets of $\Pi\left(X_{8} \frown X_{1}\right)$. The theorem is a consequence of (17), (16), and (13).
Let $n$ be a non zero natural number and $X$ be a non-empty, $n$-element finite sequence.

A family of fields of $X$ is an $n$-element finite sequence and is defined by
(Def. 3) for every natural number $i$ such that $i \in \operatorname{Seg} n$ holds $i t(i)$ is a field of subsets of $X(i)$.

Let $S$ be a family of fields of $X$ and $i$ be a natural number. Assume $i \in \operatorname{Seg} n$. Observe that the functor $S(i)$ yields a field of subsets of $X(i)$.

Observe that a family of fields of $X$ is a family of semialgebras of $X$.
Let us consider a non-empty, 1-element finite sequence $X$ and a family $S$ of fields of $X$. Now we state the propositions:
(19) the set of all $\Pi\langle s\rangle$ where $s$ is an element of $S(1)$ is a field of subsets of the set of all $\langle x\rangle$ where $x$ is an element of $X(1)$. The theorem is a consequence of (14).
(20) SemiringProduct $(S)$ is a field of subsets of $\Pi X$. The theorem is a consequence of (19).
Let $n$ be a non zero natural number, $X$ be a non-empty, $n$-element finite sequence, and $S$ be a family of fields of $X$.

A family of measures of $S$ is an $n$-element finite sequence and is defined by
(Def. 4) for every natural number $i$ such that $i \in \operatorname{Seg} n$ holds $i t(i)$ is a measure on $S(i)$.

## 3. Product of Two Measures

Let $X_{1}, X_{2}$ be sets, $S_{1}$ be a field of subsets of $X_{1}$, and $S_{2}$ be a field of subsets of $X_{2}$. The functor MeasRect $\left(S_{1}, S_{2}\right)$ yielding a semialgebra of sets of $X_{1} \times X_{2}$ is defined by the term
(Def. 5) the set of all $A \times B$ where $A$ is an element of $S_{1}, B$ is an element of $S_{2}$. Now we state the proposition:
(21) Let us consider a set $X$, and a field $F$ of subsets of $X$. Then there exists a semialgebra $S$ of sets of $X$ such that
(i) $F=S$, and
(ii) $F=$ the field generated by $S$.

Let $X_{1}, X_{2}$ be sets, $S_{1}$ be a field of subsets of $X_{1}, S_{2}$ be a field of subsets of $X_{2}, m_{1}$ be a measure on $S_{1}$, and $m_{2}$ be a measure on $S_{2}$. The functor ProdpreMeas $\left(m_{1}, m_{2}\right)$ yielding a non-negative, zeroed function from MeasRect $\left(S_{1}, S_{2}\right)$ into $\overline{\mathbb{R}}$ is defined by
(Def. 6) for every element $C$ of $\operatorname{MeasRect}\left(S_{1}, S_{2}\right)$, there exists an element $A$ of $S_{1}$ and there exists an element $B$ of $S_{2}$ such that $C=A \times B$ and $i t(C)=$ $m_{1}(A) \cdot m_{2}(B)$.
Now we state the propositions:
(22) Let us consider sets $X_{1}, X_{2}$, a field $S_{1}$ of subsets of $X_{1}$, a field $S_{2}$ of subsets of $X_{2}$, a measure $m_{1}$ on $S_{1}$, a measure $m_{2}$ on $S_{2}$, and sets $A, B$.

Suppose $A \in S_{1}$ and $B \in S_{2}$. Then (ProdpreMeas $\left.\left(m_{1}, m_{2}\right)\right)(A \times B)=$ $m_{1}(A) \cdot m_{2}(B)$.
(23) Let us consider sets $X_{1}, X_{2}$, a non empty family $S_{1}$ of subsets of $X_{1}$, a non empty family $S_{2}$ of subsets of $X_{2}$, a non empty family $S$ of subsets of $X_{1} \times X_{2}$, and a finite sequence $H$ of elements of $S$. Suppose $S=$ the set of all $A \times B$ where $A$ is an element of $S_{1}, B$ is an element of $S_{2}$. Then there exists a finite sequence $F$ of elements of $S_{1}$ and there exists a finite sequence $G$ of elements of $S_{2}$ such that len $H=\operatorname{len} F$ and len $H=$ len $G$ and for every natural number $k$ such that $k \in \operatorname{dom} H$ and $H(k) \neq \emptyset$ holds $H(k)=F(k) \times G(k)$.
Proof: For every natural number $k$ such that $k \in$ dom $H$ there exists an element $A$ of $S_{1}$ and there exists an element $B$ of $S_{2}$ such that $H(k)=$ $A \times B$. Define $\mathcal{P}$ [natural number, set] $\equiv$ there exists an element $B$ of $S_{2}$ such that $H\left(\$_{1}\right)=\$_{2} \times B$. Consider $F$ being a finite sequence of elements of $S_{1}$ such that $\operatorname{dom} F=$ Seg len $H$ and for every natural number $k$ such that $k \in \operatorname{Seg}$ len $H$ holds $\mathcal{P}[k, F(k)$ ] from [4, Sch. 5]. Define $\mathcal{Q}[$ natural number, set] $\equiv$ there exists an element $A$ of $S_{1}$ such that $H\left(\$_{1}\right)=A \times \$_{2}$. For every natural number $k$ such that $k \in$ Seg len $H$ there exists an element $B$ of $S_{2}$ such that $\mathcal{Q}[k, B]$. Consider $G$ being a finite sequence of elements of $S_{2}$ such that dom $G=$ Seg len $H$ and for every natural number $k$ such that $k \in \operatorname{Seg}$ len $H$ holds $\mathcal{Q}[k, G(k)$ ] from [4, Sch. 5].
(24) Let us consider a set $X$, a non empty, semi-diff-closed, $\cap$-closed family $S$ of subsets of $X$, and elements $E_{1}, E_{2}$ of $S$. Then there exist disjoint valued finite sequences $O, T, F$ of elements of $S$ such that
(i) $\cup \operatorname{rng} O=E_{1} \backslash E_{2}$, and
(ii) $\bigcup \operatorname{rng} T=E_{2} \backslash E_{1}$, and
(iii) $\bigcup \operatorname{rng} F=E_{1} \cap E_{2}$, and
(iv) $\left(O^{\wedge} T\right)^{\wedge} F$ is a disjoint valued finite sequence of elements of $S$.

The theorem is a consequence of (9).
(25) Let us consider sets $X_{1}, X_{2}$, a field $S_{1}$ of subsets of $X_{1}$, a field $S_{2}$ of subsets of $X_{2}$, a measure $m_{1}$ on $S_{1}$, a measure $m_{2}$ on $S_{2}$, and elements $E_{1}, E_{2}$ of MeasRect $\left(S_{1}, S_{2}\right)$. Suppose $E_{1}$ misses $E_{2}$ and $E_{1} \cup E_{2} \in$ $\operatorname{MeasRect}\left(S_{1}, S_{2}\right)$. Then (ProdpreMeas $\left.\left(m_{1}, m_{2}\right)\right)\left(E_{1} \cup E_{2}\right)=$
$\left(\operatorname{ProdpreMeas}\left(m_{1}, m_{2}\right)\right)\left(E_{1}\right)+\left(\operatorname{ProdpreMeas}\left(m_{1}, m_{2}\right)\right)\left(E_{2}\right)$. The theorem is a consequence of (1) and (22).
(26) Let us consider a non empty set $X$, a non empty family $S$ of subsets of $X$, a function $f$ from $\mathbb{N}$ into $S$, and a sequence $F$ of partial functions from $X$ into $\overline{\mathbb{R}}$. Suppose $f$ is disjoint valued and for every natural number
$n, F(n)=\chi_{f(n), X}$. Let us consider an object $x$. Suppose $x \in X$. Then $\chi_{\bigcup f, X}(x)=\left(\lim \left(\sum_{\alpha=0}^{\kappa} F(\alpha)\right)_{\kappa \in \mathbb{N}}\right)(x)$.
(27) Let us consider a non empty set $X$, a $\sigma$-field $S$ of subsets of $X$, a $\sigma$ measure $M$ on $S$, a partial function $f$ from $X$ to $\overline{\mathbb{R}}$, and a real number $r$. Suppose $\operatorname{dom} f \in S$ and $0 \leqslant r$ and for every object $x$ such that $x \in \operatorname{dom} f$ holds $f(x)=r$. Then $\int f \mathrm{~d} M=r \cdot M(\operatorname{dom} f)$.
Let us consider a non empty set $X$, a $\sigma$-field $S$ of subsets of $X$, a $\sigma$-measure $M$ on $S$, a partial function $f$ from $X$ to $\overline{\mathbb{R}}$, and an element $A$ of $S$. Now we state the propositions:
(28) Suppose there exists an element $E$ of $S$ such that $E=\operatorname{dom} f$ and $f$ is measurable on $E$ and for every object $x$ such that $x \in \operatorname{dom} f \backslash A$ holds $f(x)=0$ and $f$ is non-negative. Then $\int f \mathrm{~d} M=\int f \upharpoonright A \mathrm{~d} M$. The theorem is a consequence of (27).
(29) If $f$ is integrable on $M$ and for every object $x$ such that $x \in \operatorname{dom} f \backslash A$ holds $f(x)=0$, then $\int f \mathrm{~d} M=\int f\lceil A \mathrm{~d} M$. The theorem is a consequence of (27).
(30) Let us consider non empty sets $X_{1}, X_{2}$, a $\sigma$-field $S_{1}$ of subsets of $X_{1}$, a $\sigma$ field $S_{2}$ of subsets of $X_{2}$, a $\sigma$-measure $M_{2}$ on $S_{2}$, a function $D$ from $\mathbb{N}$ into $S_{1}$, a function $E$ from $\mathbb{N}$ into $S_{2}$, an element $A$ of $S_{1}$, an element $B$ of $S_{2}$, a sequence $F$ of partial functions from $X_{2}$ into $\overline{\mathbb{R}}$, a sequence $R$ of $\mathbb{R}^{X_{1}}$, and an element $x$ of $X_{1}$. Suppose for every natural number $n, R(n)=\chi_{D(n), X_{1}}$ and for every natural number $n, F(n)=R(n)(x) \cdot \chi_{E(n), X_{2}}$ and for every natural number $n, E(n) \subseteq B$. Then there exists a sequence $I$ of extended reals such that
(i) for every natural number $n, I(n)=M_{2}(E(n)) \cdot \chi_{D(n), X_{1}}(x)$, and
(ii) $I$ is summable, and
(iii) $\int \lim \left(\sum_{\alpha=0}^{\kappa} F(\alpha)\right)_{\kappa \in \mathbb{N}} \mathrm{d} M_{2}=\sum I$.

Proof: For every natural number $n, \operatorname{dom}(F(n))=X_{2}$. Reconsider $S_{3}=$ $X_{2}$ as an element of $S_{2}$. For every natural number $n$ and for every set $y$ such that $y \in E(n)$ holds $F(n)(y)=0$ or $F(n)(y)=1$ by [10, (3)], [18, (1)], [12, (39)]. For every natural number $n$ and for every set $y$ such that $y \notin E(n)$ holds $F(n)(y)=0$. For every natural number $n, F(n)$ is nonnegative and $F(n)$ is measurable on $B$ by [8, (51)], [17, (37)], [18, (29)]. For every element $y$ of $X_{2}$ such that $y \in B$ holds $F \# y$ is summable by [8, (51), (39)], [19, (16)], [29, (37)].

Consider $I$ being a sequence of extended reals such that for every natural number $n, I(n)=\int F(n) \upharpoonright B \mathrm{~d} M_{2}$ and $I$ is summable and $\int \lim \left(\sum_{\alpha=0}^{\kappa} F(\alpha)\right)_{\kappa \in \mathbb{N}} \backslash B \mathrm{~d} M_{2}=\sum I$. For every natural number $n, I(n)=$
$M_{2}(E(n)) \cdot \chi_{D(n), X_{1}}(x)$ by [28, (61)], [10, (47), (49)], [18, (29)]. For every natural number $n, F(n)$ is measurable on $S_{3}$ by [18, (29)], [17, (37)]. For every natural number $n, F(n)$ is without $-\infty$. For every element $y$ of $X_{2}$ such that $y \in S_{3}$ holds $\left(\sum_{\alpha=0}^{\kappa} F(\alpha)\right)_{\kappa \in \mathbb{N}} \# y$ is convergent by [19, (38)]. For every object $y$ such that $y \in \operatorname{dom} \lim \left(\sum_{\alpha=0}^{\kappa} F(\alpha)\right)_{\kappa \in \mathbb{N}} \backslash B$ holds $\left(\lim \left(\sum_{\alpha=0}^{\kappa} F(\alpha)\right)_{\kappa \in \mathbb{N}}\right)(y)=0$ by [19, (43)], [16, (52)]. For every object $y$ such that $y \in \operatorname{dom} \lim \left(\sum_{\alpha=0}^{\kappa} F(\alpha)\right)_{\kappa \in \mathbb{N}}$ holds $\left(\lim \left(\sum_{\alpha=0}^{\kappa} F(\alpha)\right)_{\kappa \in \mathbb{N}}\right)(y) \geqslant 0$ by [19, (36)], [8, (51)], [19, (10), (38)].
(31) Let us consider a non empty set $X$, a $\sigma$-field $S$ of subsets of $X$, an element $A$ of $S$, and an extended real number $p$. Then $X \longmapsto p$ is measurable on $A$. Proof: For every real number $r, A \cap \operatorname{GTE}-\operatorname{dom}(X \longmapsto p, r) \in S$ by [26, (7)], [7, (7)].

Let $A, X$ be sets. The functor $\bar{\chi}_{A, X}$ yielding a function from $X$ into $\overline{\mathbb{R}}$ is defined by
(Def. 7) for every object $x$ such that $x \in X$ holds if $x \in A$, then $i t(x)=+\infty$ and if $x \notin A$, then $i t(x)=0$.
Now we state the proposition:
(32) Let us consider a non empty set $X$, a $\sigma$-field $S$ of subsets of $X$, and elements $A, B$ of $S$. Then $\bar{\chi}_{A, X}$ is measurable on $B$.
Let $X, A$ be sets. Let us observe that $\bar{\chi}_{A, X}$ is non-negative.
(33) Let us consider a non empty set $X$, a $\sigma$-field $S$ of subsets of $X$, a $\sigma$ measure $M$ on $S$, and an element $A$ of $S$. Then
(i) if $M(A) \neq 0$, then $\int \bar{\chi}_{A, X} \mathrm{~d} M=+\infty$, and
(ii) if $M(A)=0$, then $\int \bar{\chi}_{A, X} \mathrm{~d} M=0$.

Proof: Reconsider $X_{3}=X$ as an element of $S$. Reconsider $X_{2}=X_{3} \backslash A$ as an element of $S$. Reconsider $F=\bar{\chi}_{A, X} \upharpoonright A$ as a partial function from $X$ to $\overline{\mathbb{R}}$. Reconsider $O=\bar{\chi}_{A, X} \upharpoonright X_{2}$ as a partial function from $X$ to $\overline{\mathbb{R}}$. Reconsider $T=\bar{\chi}_{A, X} \upharpoonright\left(X_{2} \cup A\right)$ as a partial function from $X$ to $\overline{\mathbb{R}} . \int F \mathrm{~d} M=0 . O$ is measurable on $X_{2}$. For every element $x$ of $X$ such that $x \in \operatorname{dom}\left(\bar{\chi}_{A, X} \upharpoonright X_{2}\right)$ holds $\left(\bar{\chi}_{A, X} \upharpoonright X_{2}\right)(x)=0$ by [10, (47)]. $\int T \mathrm{~d} M=\int O \mathrm{~d} M+0$.
(34) Let us consider non empty sets $X_{1}, X_{2}$, a $\sigma$-field $S_{1}$ of subsets of $X_{1}$, a $\sigma$-field $S_{2}$ of subsets of $X_{2}$, a $\sigma$-measure $M_{1}$ on $S_{1}$, a $\sigma$-measure $M_{2}$ on $S_{2}$, and a disjoint valued function $K$ from $\mathbb{N}$ into $\operatorname{MeasRect}\left(S_{1}, S_{2}\right)$. Suppose $\bigcup K \in \operatorname{MeasRect}\left(S_{1}, S_{2}\right)$. Then (ProdpreMeas $\left.\left(M_{1}, M_{2}\right)\right)(\bigcup K)=$ $\bar{\sum}\left(\right.$ ProdpreMeas $\left.\left(M_{1}, M_{2}\right) \cdot K\right)$.
Proof: Consider $A$ being an element of $S_{1}, B$ being an element of $S_{2}$ such that $\cup K=A \times B$. Consider $P$ being an element of $S_{1}, Q$ being an element of $S_{2}$ such that $\cup K=P \times Q$ and (ProdpreMeas $\left.\left(M_{1}, M_{2}\right)\right)(\bigcup K)=M_{1}(P)$.
$M_{2}(Q)$. Define $\mathcal{F}($ object $)=\chi_{K\left(\$_{1}\right), X_{1} \times X_{2}}$. Consider $X_{6}$ being a sequence of partial functions from $X_{1} \times X_{2}$ into $\mathbb{R}$ such that for every natural number $n, X_{6}(n)=\mathcal{F}(n)$ from [24, Sch. 1]. Define $\mathcal{P}$ [natural number, object] $\equiv$ $\$_{2}=\pi_{1}\left(K\left(\$_{1}\right)\right)$. For every element $i$ of $\mathbb{N}$, there exists an element $A$ of $S_{1}$ such that $\mathcal{P}[i, A]$ by [2, (9)], [7, (7)]. Consider $D$ being a function from $\mathbb{N}$ into $S_{1}$ such that for every element $i$ of $\mathbb{N}, \mathcal{P}[i, D(i)]$ from [11, Sch. 3]. Define $\mathcal{Q}$ [natural number, object] $\equiv \$_{2}=\pi_{2}\left(K\left(\$_{1}\right)\right)$. For every element $i$ of $\mathbb{N}$, there exists an element $B$ of $S_{2}$ such that $\mathcal{Q}[i, B]$ by [2, (9)], [7, (7)].

Consider $E$ being a function from $\mathbb{N}$ into $S_{2}$ such that for every element $i$ of $\mathbb{N}, \mathcal{Q}[i, E(i)]$ from [11, Sch. 3]. Define $\mathcal{O}($ object $)=\chi_{D\left(\$_{1}\right), X_{1}}$. Consider $X_{7}$ being a sequence of partial functions from $X_{1}$ into $\overline{\mathbb{R}}$ such that for every natural number $n, X_{7}(n)=\mathcal{O}(n)$ from [24, Sch. 1]. Define $\mathcal{T}$ (object) $=\chi_{E\left(\$_{1}\right), X_{2}}$. Consider $X_{4}$ being a sequence of partial functions from $X_{2}$ into $\overline{\mathbb{R}}$ such that for every natural number $n, X_{4}(n)=\mathcal{T}(n)$ from [24, Sch. 1]. For every natural number $n$ and for every objects $x, y$ such that $x \in X_{1}$ and $y \in X_{2}$ holds $X_{6}(n)(x, y)=X_{7}(n)(x) \cdot X_{4}(n)(y)$ by [14, (87)], [2, (9)], (2). (ProdpreMeas $\left.\left(M_{1}, M_{2}\right)\right)(\cup K)=M_{1}(A) \cdot M_{2}(B)$ by [14, (110)]. Reconsider $C_{1}=\chi_{A \times B, X_{1} \times X_{2}}$ as a function from $X_{1} \times X_{2}$ into $\overline{\mathbb{R}}$. For every element $x$ of $X_{1}, M_{2}(B) \cdot \chi_{A, X_{1}}(x)=\int \operatorname{curry}\left(C_{1}, x\right) \mathrm{d} M_{2}$ by (2), [13, (5)], [19, (14)], [23, (4)]. For every object $n$ such that $n \in \mathbb{N}$ holds $X_{7}(n) \in \mathbb{R}^{X_{1}}$ by [12, (39)]. Reconsider $R_{1}=X_{7}$ as a sequence of $\mathbb{R}^{X_{1}}$. For every natural number $n, D(n) \subseteq A$ and $E(n) \subseteq B$ by [2, (10)], [1, (1)]. For every element $x$ of $X_{1}$, there exists a sequence $X_{5}$ of partial functions from $X_{2}$ into $\overline{\mathbb{R}}$ and there exists a sequence $I$ of extended reals such that for every natural number $n, X_{5}(n)=R_{1}(n)(x) \cdot \chi_{E(n), X_{2}}$ and for every natural number $n, I(n)=M_{2}(E(n)) \cdot \chi_{D(n), X_{1}}(x)$ and $I$ is summable and $\int \lim \left(\sum_{\alpha=0}^{\kappa} X_{5}(\alpha)\right)_{\kappa \in \mathbb{N}} \mathrm{d} M_{2}=\sum I$ by [13, (45)], (30).

Reconsider $L_{1}=\lim \left(\sum_{\alpha=0}^{\kappa} X_{6}(\alpha)\right)_{\kappa \in \mathbb{N}}$ as a function from $X_{1} \times$ $X_{2}$ into $\overline{\mathbb{R}}$. For every element $x$ of $X_{1}$ and for every element $y$ of $X_{2}$, $\left(\operatorname{curry}\left(C_{1}, x\right)\right)(y)=\left(\operatorname{curry}\left(L_{1}, x\right)\right)(y)$. For every element $x$ of $X_{1}, \operatorname{curry}\left(C_{1}\right.$, $x)=\operatorname{curry}\left(L_{1}, x\right)$. For every element $x$ of $X_{1}, M_{2}(B) \cdot \chi_{A, X_{1}}(x)=\int \operatorname{curry}$ $\left(L_{1}, x\right) \mathrm{d} M_{2}$. For every element $x$ of $X_{1}$, there exists a sequence $I$ of extended reals such that for every natural number $n, I(n)=M_{2}(E(n))$. $\chi_{D(n), X_{1}}(x)$ and $M_{2}(B) \cdot \chi_{A, X_{1}}(x)=\sum I$ by [8, (51)], [19, (38), (29), (30)]. Define $\mathcal{R}\left[\right.$ natural number, object] $\equiv$ if $M_{2}\left(E\left(\$_{1}\right)\right)=+\infty$, then $\$_{2}=\bar{\chi}_{D\left(\$_{1}\right), X_{1}}$ and if $M_{2}\left(E\left(\$_{1}\right)\right) \neq+\infty$, then there exists a real number $m_{2}$ such that $m_{2}=M_{2}\left(E\left(\$_{1}\right)\right)$ and $\$_{2}=m_{2} \cdot \chi_{D\left(\$_{1}\right), X_{1}}$. For every element $n$ of $\mathbb{N}$, there exists an element $y$ of $X_{1} \rightarrow \overline{\mathbb{R}}$ such that $\mathcal{R}[n, y]$ by [13, (45)], [8, (51)]. Consider $F_{1}$ being a function from $\mathbb{N}$ into $X_{1} \dot{\rightarrow} \overline{\mathbb{R}}$ such that for every element $n$ of $\mathbb{N}, \mathcal{R}\left[n, F_{1}(n)\right]$ from [11, Sch. 3]. For every natural number
$n$, $\operatorname{dom}\left(F_{1}(n)\right)=X_{1}$. For every natural number $n, F_{1}(n)$ is non-negative by [8, (51)]. For every natural numbers $n, m, \operatorname{dom}\left(F_{1}(n)\right)=\operatorname{dom}\left(F_{1}(m)\right)$.

Reconsider $X_{3}=X_{1}$ as an element of $S_{1}$. For every natural number $n, F_{1}(n)$ is non-negative and $F_{1}(n)$ is measurable on $A$ and $F_{1}(n)$ is measurable on $X_{3}$ by (32), [18, (29)], [17, (37)]. For every element $x$ of $X_{1}$ such that $x \in A$ holds $F_{1} \# x$ is summable by [ 8 , (51), (39)], [20, (2)]. Consider $J$ being a sequence of extended reals such that for every natural number $n, J(n)=\int F_{1}(n) \upharpoonright A \mathrm{~d} M_{1}$ and $J$ is summable and $\int \lim \left(\sum_{\alpha=0}^{\kappa} F_{1}(\alpha)\right)_{\kappa \in \mathbb{N}} \upharpoonright A \mathrm{~d} M_{1}=\sum J$. For every natural number $n, J(n)=$ $\int F_{1}(n) \mathrm{d} M_{1}$. Reconsider $X_{3}=X_{1}$ as an element of $S_{1}$. For every element $n$ of $\mathbb{N}, J(n)=\left(\operatorname{ProdpreMeas}\left(M_{1}, M_{2}\right) \cdot K\right)(n)$ by $(33),[8,(51)],[18,(29)]$, [16, (86), (88)]. For every element $x$ of $X_{1},\left(\lim \left(\sum_{\alpha=0}^{\kappa} F_{1}(\alpha)\right)_{\kappa \in \mathbb{N}}\right)(x) \geqslant 0$ by [19, (38)], [29, (37), (23)], [8, (51)]. For every natural number $n, F_{1}(n)$ is measurable on $X_{3}$ and $F_{1}(n)$ is without $-\infty$. For every object $x$ such that $x \in \operatorname{dom} \lim \left(\sum_{\alpha=0}^{\kappa} F_{1}(\alpha)\right)_{\kappa \in \mathbb{N}} \backslash A$ holds $\left(\lim \left(\sum_{\alpha=0}^{\kappa} F_{1}(\alpha)\right)_{\kappa \in \mathbb{N}}\right)(x)=0$ by [19, (30), (32)], [16, (52)]. $\int \lim \left(\sum_{\alpha=0}^{\kappa} F_{1}(\alpha)\right)_{\kappa \in \mathbb{N}} \mathrm{d} M_{1}=$ $\int \lim \left(\sum_{\alpha=0}^{\kappa} F_{1}(\alpha)\right)_{\kappa \in \mathbb{N}} \upharpoonright A \mathrm{~d} M_{1} . \int \lim \left(\sum_{\alpha=0}^{\kappa} F_{1}(\alpha)\right)_{\kappa \in \mathbb{N}} \mathrm{d} M_{1}=M_{1}(A) \cdot M_{2}(B)$ by [11, (63)], [19, (30), (32)], [8, (51)].
(35) Let us consider a without $-\infty$ finite sequence $f$ of elements of $\overline{\mathbb{R}}$, and a without $-\infty$ sequence $s$ of extended reals. Suppose for every object $n$ such that $n \in \operatorname{dom} f$ holds $f(n)=s(n)$.
Then $\sum f+s(0)=\left(\sum_{\alpha=0}^{\kappa} s(\alpha)\right)_{\kappa \in \mathbb{N}}(\operatorname{len} f)$.
Proof: Consider $F$ being a sequence of $\overline{\mathbb{R}}$ such that $\sum f=F(\operatorname{len} f)$ and $F(0)=0$ and for every natural number $i$ such that $i<\operatorname{len} f$ holds $F(i+1)=F(i)+f(i+1)$. Define $\mathcal{P}$ [natural number] $\equiv$ if $\$_{1} \leqslant \operatorname{len} f$, then $F\left(\$_{1}\right)+s(0)=\left(\sum_{\alpha=0}^{\kappa} s(\alpha)\right)_{\kappa \in \mathbb{N}}\left(\$_{1}\right)$ and $F\left(\$_{1}\right) \neq-\infty$. For every natural number $k$ such that $\mathcal{P}[k]$ holds $\mathcal{P}[k+1$ ] by [3, (11)], [27, (25)], [16, (10)], [3, (13)]. For every natural number $k, \mathcal{P}[k$ ] from [3, Sch. 2].
(36) Let us consider a non-negative finite sequence $f$ of elements of $\overline{\mathbb{R}}$, and a sequence $s$ of extended reals. Suppose for every object $n$ such that $n \in$ dom $f$ holds $f(n)=s(n)$ and for every element $n$ of $\mathbb{N}$ such that $n \notin \operatorname{dom} f$ holds $s(n)=0$. Then
(i) $\sum f=\sum s$, and
(ii) $\sum f=\bar{\sum} s$.

Proof: For every object $n$ such that $n \in \operatorname{dom} s$ holds $0 \leqslant s(n)$ by [8, (51)]. $\sum f+s(0)=\left(\sum_{\alpha=0}^{\kappa} s(\alpha)\right)_{\kappa \in \mathbb{N}}($ len $f)$. Define $\mathcal{P}$ [natural number] $\equiv$ $\left(\sum_{\alpha=0}^{\kappa} s(\alpha)\right)_{\kappa \in \mathbb{N}}(\operatorname{len} f)=\left(\left(\sum_{\alpha=0}^{\kappa} s(\alpha)\right)_{\kappa \in \mathbb{N}} \uparrow\right.$ len $\left.f\right)\left(\$_{1}\right)$. For every natural number $k$ such that $\mathcal{P}[k]$ holds $\mathcal{P}[k+1]$ by [27, (25)]. For every natural number $k, \mathcal{P}[k]$ from [3, Sch. 2].
(37) Let us consider non empty sets $X_{1}, X_{2}$, a $\sigma$-field $S_{1}$ of subsets of $X_{1}$, a $\sigma$ field $S_{2}$ of subsets of $X_{2}$, a $\sigma$-measure $M_{1}$ on $S_{1}$, a $\sigma$-measure $M_{2}$ on $S_{2}$, and a disjoint valued finite sequence $F$ of elements of $\operatorname{MeasRect}\left(S_{1}, S_{2}\right)$. Suppose $\bigcup F \in \operatorname{MeasRect}\left(S_{1}, S_{2}\right)$. Then (ProdpreMeas $\left.\left(M_{1}, M_{2}\right)\right)(\bigcup F)=$ $\sum\left(\right.$ ProdpreMeas $\left.\left(M_{1}, M_{2}\right) \cdot F\right)$.
Proof: Set $S=\operatorname{MeasRect}\left(S_{1}, S_{2}\right)$. Define $\mathcal{P}\left[\right.$ object, object] $\equiv$ if $\$_{1} \in$ $\operatorname{dom} F$, then $\$_{2}=F\left(\$_{1}\right)$ and if $\$_{1} \notin \operatorname{dom} F$, then $\$_{2}=\emptyset$. For every element $n$ of $\mathbb{N}$, there exists an element $y$ of $S$ such that $\mathcal{P}[n, y]$ by [10, (3)]. Consider $G$ being a function from $\mathbb{N}$ into $S$ such that for every element $n$ of $\mathbb{N}, \mathcal{P}[n, G(n)]$ from [11, Sch. 3]. For every object $x$ such that $x \notin$ dom $F$ holds $G(x)=\emptyset$. For every objects $x, y$ such that $x \neq y$ holds $G(x)$ misses $G(y)$. (ProdpreMeas $\left.\left(M_{1}, M_{2}\right)\right)(\bigcup F)=\bar{\sum}\left(\operatorname{ProdpreMeas}\left(M_{1}, M_{2}\right)\right.$. $G)$. For every object $n$ such that $n \in \operatorname{dom}\left(\operatorname{ProdpreMeas}\left(M_{1}, M_{2}\right) \cdot F\right)$ holds $\left(\operatorname{ProdpreMeas}\left(M_{1}, M_{2}\right) \cdot F\right)(n)=\left(\operatorname{ProdpreMeas}\left(M_{1}, M_{2}\right) \cdot G\right)(n)$ by [10, (11), (12), (13)]. For every element $n$ of $\mathbb{N}$ such that $n \notin \operatorname{dom}$ (ProdpreMeas $\left.\left(M_{1}, M_{2}\right) \cdot F\right)$ holds (ProdpreMeas $\left.\left(M_{1}, M_{2}\right) \cdot G\right)(n)=0$ by [10, (3), (11), (13)].
(38) Let us consider non empty sets $X_{1}, X_{2}$, a $\sigma$-field $S_{1}$ of subsets of $X_{1}$, a $\sigma$-field $S_{2}$ of subsets of $X_{2}$, a $\sigma$-measure $M_{1}$ on $S_{1}$, and a $\sigma$-measure $M_{2}$ on $S_{2}$. Then ProdpreMeas $\left(M_{1}, M_{2}\right)$ is a pre-measure of $\operatorname{MeasRect}\left(S_{1}, S_{2}\right)$. The theorem is a consequence of (37) and (34).
Let $X_{1}, X_{2}$ be non empty sets, $S_{1}$ be a $\sigma$-field of subsets of $X_{1}, S_{2}$ be a $\sigma$ field of subsets of $X_{2}, M_{1}$ be a $\sigma$-measure on $S_{1}$, and $M_{2}$ be a $\sigma$-measure on $S_{2}$. Let us observe that the functor $\operatorname{ProdpreMeas}\left(M_{1}, M_{2}\right)$ yields a pre-measure of $\operatorname{MeasRect}\left(S_{1}, S_{2}\right)$.

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# Conservation Rules of Direct Sum Decomposition of Groups 

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#### Abstract

Summary. In this article, conservation rules of the direct sum decomposition of groups are mainly discussed. In the first section, we prepare miscellaneous definitions and theorems for further formalization in Mizar [5]. In the next three sections, we formalized the fact that the property of direct sum decomposition is preserved against the substitutions of the subscript set, flattening of direct sum, and layering of direct sum, respectively. We referred to [14, [13] [6] and 11] in the formalization.


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## 1. Preliminaries

Let $I, J$ be non empty sets, $a$ be a function from $I$ into $J$, and $F$ be a multiplicative magma family of $J$. Observe that the functor $F \cdot a$ yields a multiplicative magma family of $I$. Let $F$ be a group family of $J$. Let us observe that the functor $F \cdot a$ yields a group family of $I$. Let $G$ be a group and $F$ be a subgroup family of $J$ and $G$. The functor $F \cdot a$ yielding a subgroup family of $I$ and $G$ is defined by the term
(Def. 1) $F \cdot a$.

The scheme $S c h 1$ deals with a set $\mathcal{A}$ and a 1 -sorted structure $\mathcal{B}$ and a unary functor $\mathcal{F}$ yielding a set and states that
(Sch. 1) There exists a function $f$ such that $\operatorname{dom} f=\mathcal{A}$ and for every element $x$ of $\mathcal{B}$ such that $x \in \mathcal{A}$ holds $f(x)=\mathcal{F}(x)$.
Let $I$ be a set. Let us note that there exists a many sorted set indexed by $I$ which is non-empty and disjoint valued.

Now we state the propositions:
(1) Let us consider a non-empty, disjoint valued function $f$. If $\bigcup f$ is finite, then $\operatorname{dom} f$ is finite.
Proof: For every objects $x, y$ such that $x, y \in \operatorname{dom} f$ and $f(x)=f(y)$ holds $x=y$ by [7, (3)].
(2) Let us consider non empty sets $X, Y$, sets $X_{0}, Y_{0}$, and a function $f$ from $X$ into $Y$. Suppose $f$ is bijective and $\operatorname{rng}\left(f \upharpoonright X_{0}\right)=Y_{0}$. Then $\left(f \upharpoonright X_{0}\right)^{-1}=$ $f^{-1} \upharpoonright Y_{0}$.
Proof: For every object $x$ such that $x \in \operatorname{dom}\left(f^{-1} \upharpoonright Y_{0}\right)$ holds $\left(f^{-1} \mid Y_{0}\right)(x)=$ $\left(f \upharpoonright X_{0}\right)^{-1}(x)$ by [18, (62)], [7, (49), (33)], [18, (59)].

## 2. Conservation Rule of Direct Sum Decomposition for Substitution of Subscript Set

Now we state the proposition:
(3) Let us consider non empty sets $I, J$, a function $a$ from $I$ into $J$, a multiplicative magma family $F$ of $J$, and an element $x$ of $\Pi F$. Then $x \cdot a \in \Pi(F \cdot a)$. Proof: Reconsider $y=x \cdot a$ as a many sorted set indexed by $I$. Reconsider $z=$ the support of $F \cdot a$ as a many sorted set indexed by $I$. For every object $i$ such that $i \in I$ holds $y(i) \in z(i)$ by [7, (13)].
Let $I, J$ be non empty sets, $a$ be a function from $I$ into $J$, and $F$ be a multiplicative magma family of $J$. The functor $\operatorname{Trans} \prod(F, a)$ yielding a function from $\prod F$ into $\prod(F \cdot a)$ is defined by
(Def. 2) for every element $x$ of $\Pi F, i t(x)=x \cdot a$.
Now we state the proposition:
(4) Let us consider non empty sets $I, J$, a function $a$ from $I$ into $J$, and a multiplicative magma family $F$ of $J$. Then $\operatorname{Trans} \prod(F, a)$ is multiplicative.
Proof: Reconsider $f=\operatorname{Trans} \prod(F, a)$ as a function from $\prod F$ into $\prod(F \cdot a)$. For every elements $x, y$ of $\Pi F, f(x \cdot y)=f(x) \cdot f(y)$ by (3), [7, (13)], [10, (1)], [18, (27)].

Let $I, J$ be non empty sets, $a$ be a function from $I$ into $J$, and $F$ be a group family of $J$. Let us observe that the functor $\operatorname{Trans} \prod(F, a)$ yields a homomorphism from $\Pi F$ to $\Pi(F \cdot a)$. Now we state the propositions:
(5) Let us consider non empty sets $I, J$, a function $a$ from $I$ into $J$, a multiplicative magma family $F$ of $J$, and an element $y$ of $\Pi(F \cdot a)$. If $a$ is bijective, then $y \cdot a^{-1} \in \Pi F$.
Proof: Set $x=y \cdot a^{-1}$. For every object $j$ such that $j \in J$ holds $x(j) \in$ (the support of $F)(j)$ by [7, (32), (13)].
(6) Let us consider non empty sets $I, J$, a function $a$ from $I$ into $J$, and functions $x, y$. Suppose $\operatorname{dom} x=I$ and $\operatorname{dom} y=J$ and $a$ is bijective. Then $x=y \cdot a$ if and only if $y=x \cdot a^{-1}$.
(7) Let us consider non empty sets $I, J$, a multiplicative magma family $F$ of $J$, and a function $a$ from $I$ into $J$. Suppose $a$ is bijective. Then
(i) dom Trans $\prod(F, a)=\Omega_{\prod F}$, and
(ii) $\operatorname{rng} \operatorname{Trans} \Pi(F, a)=\Omega_{\prod(F \cdot a)}$.

The theorem is a consequence of (5) and (6).
(8) Let us consider non empty sets $I$, $J$, a function $a$ from $I$ into $J$, and a multiplicative magma family $F$ of $J$. If $a$ is bijective, then $\operatorname{Trans} \Pi(F, a)$ is bijective.
Proof: Reconsider $f=\operatorname{Trans} \prod(F, a)$ as a function from $\prod F$ into $\prod(F \cdot a)$. $\operatorname{dom} f=\Omega_{\prod F}$ and $\operatorname{rng} f=\Omega_{\prod(F \cdot a)}$. For every objects $x, y$ such that $x$, $y \in \operatorname{dom} f$ and $f(x)=f(y)$ holds $x=y$ by [7, (86)].
Let us consider non empty sets $I$, $J$, a function $a$ from $I$ into $J$, a group family $F$ of $J$, and a function $x$. Now we state the propositions:
(9) If $a$ is one-to-one, then $a^{\circ}(\operatorname{support}(x \cdot a, F \cdot a)) \subseteq \operatorname{support}(x, F)$.

Proof: For every object $j$ such that $j \in a^{\circ}$ (support $\left.(x \cdot a, F \cdot a)\right)$ holds $j \in \operatorname{support}(x, F)$ by [7, (13)].
(10) If $a$ is onto, then $\operatorname{support}(x, F) \subseteq a^{\circ}(\operatorname{support}(x \cdot a, F \cdot a))$.

Proof: For every object $j$ such that $j \in \operatorname{support}(x, F)$ holds $j \in a^{\circ}(\operatorname{support}(x \cdot a, F \cdot a))$ by [8, (11)], [7, (13)].
(11) If $a$ is one-to-one, then if $x \in \operatorname{sum} F$, then $x \cdot a \in \operatorname{sum}(F \cdot a)$. The theorem is a consequence of (3) and (9).
(12) If $a$ is bijective, then $x \in \operatorname{sum} F$ iff $x \cdot a \in \operatorname{sum}(F \cdot a)$ and $\operatorname{dom} x=J$. The theorem is a consequence of (11).
Let $I, J$ be non empty sets, $a$ be a function from $I$ into $J$, and $F$ be a group family of $J$. Assume $a$ is bijective. The functor $\operatorname{Trans} \sum(F, a)$ yielding a function from $\operatorname{sum} F$ into $\operatorname{sum}(F \cdot a)$ is defined by the term
(Def. 3) Trans $\prod(F, a) \upharpoonright \operatorname{sum} F$.
Now we state the proposition:
(13) Let us consider groups $G, H$, a subgroup $H_{0}$ of $H$, and a homomorphism $f$ from $G$ to $H$. Suppose $\operatorname{rng} f \subseteq \Omega_{H_{0}}$. Then $f$ is a homomorphism from $G$ to $H_{0}$.
Proof: Reconsider $g=f$ as a function from $G$ into $H_{0}$. For every elements $a, b$ of $G, g(a \cdot b)=g(a) \cdot g(b)$ by [16, (43)].
Let $I, J$ be non empty sets, $a$ be a function from $I$ into $J$, and $F$ be a group family of $J$. Assume $a$ is bijective. Let us observe that the functor $\operatorname{Trans} \sum(F, a)$ yields a homomorphism from $\operatorname{sum} F$ to $\operatorname{sum}(F \cdot a)$. Now we state the propositions:
(14) Let us consider non empty sets $I, J$, a function $a$ from $I$ into $J$, and a group family $F$ of $J$. If $a$ is bijective, then $\operatorname{Trans} \sum(F, a)$ is bijective.
Proof: Reconsider $f=\operatorname{Trans} \prod(F, a)$ as a homomorphism from $\prod F$ to $\Pi(F \cdot a)$. Reconsider $g=\operatorname{Trans} \sum(F, a)$ as a homomorphism from sum $F$ to $\operatorname{sum}(F \cdot a) . f$ is bijective. For every object $y$ such that $y \in \Omega_{\operatorname{sum}(F \cdot a)}$ holds $y \in \operatorname{rng} g$ by [16, (42)], (5), (6), (12).
(15) Let us consider a group $G$, non empty sets $I$, $J$, a direct sum components $F$ of $G$ and $J$, and a function $a$ from $I$ into $J$. If $a$ is bijective, then $F \cdot a$ is a direct sum components of $G$ and $I$. The theorem is a consequence of (14).
(16) Let us consider a non empty set $I$, and a group $G$. Then every internal direct sum components of $G$ and $I$ is a subgroup family of $I$ and $G$.
(17) Let us consider non empty sets $I, J$, a group $G$, a function $x$ from $I$ into $G$, a function $y$ from $J$ into $G$, and a function $a$ from $I$ into $J$. Suppose $a$ is onto and $x=y \cdot a$. Then support $y=a^{\circ}(\operatorname{support} x)$.
(18) Let us consider non empty sets $I, J$, a commutative group $G$, a finitesupport function $x$ from $I$ into $G$, a finite-support function $y$ from $J$ into $G$, and a function $a$ from $I$ into $J$. If $a$ is bijective and $x=y \cdot a$, then $\Pi x=\Pi y$.
Proof: Reconsider $S_{1}=\operatorname{support} x$ as a finite set. Reconsider $S_{2}=$ support $y$ as a finite set. Reconsider $s_{1}=\operatorname{CFS}\left(S_{1}\right)$ as a finite sequence of elements of $S_{1}$. Reconsider $s_{2}=\operatorname{CFS}\left(S_{2}\right)$ as a finite sequence of elements of $S_{2}$. Reconsider $x_{1}=x \upharpoonright S_{1}$ as a function from $S_{1}$ into $G$. Consider $x_{2}$ being a finite sequence of elements of $G$ such that $\prod x_{1}=\prod x_{2}$ and $x_{2}=x_{1} \cdot s_{1}$. Reconsider $y_{1}=y \upharpoonright S_{2}$ as a function from $S_{2}$ into $G$. Consider $y_{2}$ being a finite sequence of elements of $G$ such that $\prod y_{1}=\prod y_{2}$ and $y_{2}=y_{1} \cdot s_{2}$. $S_{2}=a^{\circ} S_{1} \cdot \overline{\overline{S_{1}}}=\overline{\overline{S_{2}}}$ by [1, (66)], [8, (25)], [17, (63)], [8, (17), (29)]. Reconsider $n=\overline{\overline{S_{1}}}$ as a natural number. Reconsider $a_{1}=a \upharpoonright S_{1}$ as a function from $S_{1}$ into $J$. Reconsider $a_{2}=s_{2}^{-1}$ as a function from $S_{2}$ into Seg $n$.

Reconsider $p=a_{2} \cdot a_{1} \cdot s_{1}$ as a function. If $S_{2}$ is not empty, then $x_{2}=y_{2} \cdot p$ by [18, (27)], [7, (3), (12), (47)].
(19) Let us consider non empty sets $I, J$, a group $G$, a finite-support function $x$ from $I$ into $G$, a finite-support function $y$ from $J$ into $G$, and a function $a$ from $I$ into $J$. Suppose $a$ is bijective and $x=y \cdot a$ and for every elements $i, j$ of $I, x(i) \cdot x(j)=x(j) \cdot x(i)$. Then $\Pi x=\prod y$. The theorem is a consequence of (18).
(20) Let us consider a group $G$, non empty sets $I$, $J$, an internal direct sum components $F$ of $G$ and $J$, and a function $a$ from $I$ into $J$. Suppose $a$ is bijective. Then $F \cdot a$ is an internal direct sum components of $G$ and $I$.
Proof: Reconsider $E=F \cdot a$ as a direct sum components of $G$ and $I$. For every element $i$ of $I, E(i)$ is a subgroup of $G$ by [7, (13)]. There exists a homomorphism $h$ from $\operatorname{sum} E$ to $G$ such that $h$ is bijective and for every finite-support function $x$ from $I$ into $G$ such that $x \in \operatorname{sum} E$ holds $h(x)=\prod x$ by (14), [17, (62), (63)], [12, (25)].

## 3. Conservation Rule of Direct Sum Decomposition for Flattening

Let $I$ be a non empty set and $J$ be a many sorted set indexed by $I$.
A $J$-indexed family of multiplicative magma families is a many sorted set indexed by $I$ and is defined by
(Def. 4) for every element $i$ of $I$, it $(i)$ is a multiplicative magma family of $J(i)$.
A $J$-indexed family of group families is a $J$-indexed family of multiplicative magma families and is defined by
(Def. 5) for every element $i$ of $I$, it $(i)$ is a group family of $J(i)$.
Let $N$ be a $J$-indexed family of multiplicative magma families and $i$ be an element of $I$. One can verify that the functor $N(i)$ yields a multiplicative magma family of $J(i)$. Let $N$ be a $J$-indexed family of group families. Observe that the functor $N(i)$ yields a group family of $J(i)$. Let $J$ be a disjoint valued many sorted set indexed by $I$ and $F$ be a $J$-indexed family of group families. One can verify that the functor $\bigcup F$ yields a group family of $\cup J$. Now we state the proposition:
(21) Let us consider a non empty set $I$, a disjoint valued many sorted set $J$ indexed by $I$, a $J$-indexed family of group families $F$, an element $j$ of $I$, and an object $i$. If $i \in J(j)$, then $(\bigcup F)(i)=F(j)(i)$.
Let $I$ be a non empty set, $J$ be a many sorted set indexed by $I$, and $F$ be a $J$ indexed family of multiplicative magma families. The functor ProdBundle $(F)$ yielding a multiplicative magma family of $I$ is defined by
(Def. 6) for every element $i$ of $I, i t(i)=\Pi(F(i))$.
Let $F$ be a $J$-indexed family of group families.
Note that the functor ProdBundle $(F)$ yields a group family of $I$. The functor SumBundle $(F)$ yielding a group family of $I$ is defined by
(Def. 7) for every element $i$ of $I, i t(i)=\operatorname{sum}(F(i))$.
Let $F$ be a $J$-indexed family of multiplicative magma families. The functor $\mathrm{d} \Pi F$ yielding a multiplicative magma is defined by the term
(Def. 8) $\Pi$ ProdBundle $(F)$.
Let $J$ be a non-empty many sorted set indexed by $I$. One can check that $\mathrm{d} \Pi F$ is non empty and constituted functions.

Let $F$ be a $J$-indexed family of group families. Observe that $\mathrm{d} \Pi F$ is grouplike and associative.

The functor $\mathrm{d} \sum F$ yielding a group is defined by the term
(Def. 9) sum SumBundle ( $F$ ).
Note that $\mathrm{d} \sum F$ is non empty and constituted functions.
Let us consider a non empty set $I$ and group families $F_{1}, F_{2}$ of $I$.
Let us assume that for every element $i$ of $I, F_{1}(i)$ is a subgroup of $F_{2}(i)$. Now we state the propositions:
(22) $\Pi F_{1}$ is a subgroup of $\Pi F_{2}$.

Proof: For every object $x$ such that $x \in \Omega_{\prod_{1}}$ holds $x \in \Omega_{\prod_{2}}$. Reconsider $f_{2}=$ (the multiplication of $\Pi F_{2}$ ) $\Omega_{\Gamma_{1}}$ as a function from $\Omega_{\prod F_{1}} \times \Omega_{F_{1}}$ into $\Omega_{\prod_{2}}$. Reconsider $f_{1}=$ the multiplication of $\Pi F_{1}$ as a function from $\Omega_{\prod_{1}} \times \Omega_{F_{1}}$ into $\Omega_{\prod_{2}}$. For every sets $x, y$ such that $x, y \in \Omega_{F_{1}}$ holds $f_{1}(x, y)=f_{2}(x, y)$ by [10, (1)], [16, (43)], [7, (49)], [9, (87)].
(23) $\operatorname{sum} F_{1}$ is a subgroup of $\operatorname{sum} F_{2}$.

Proof: For every object $x$ such that $x \in \Omega_{\text {sum } F_{1}}$ holds $x \in \Omega_{\text {sum } F_{2}}$ by [16, (40)], (22), [16, (42), (44)]. $\Pi F_{1}$ is a subgroup of $\Pi F_{2}$.
(24) Let us consider a non empty set $I$, a non-empty many sorted set $J$ indexed by $I$, and a $J$-indexed family of group families $F$. Then $\mathrm{d} \sum F$ is a subgroup of $\mathrm{d} \Pi F$. The theorem is a consequence of (22).
Let $I$ be a non empty set, $J$ be a non-empty, disjoint valued many sorted set indexed by $I$, and $F$ be a $J$-indexed family of group families. One can verify that the functor $\mathrm{d} \sum F$ yields a subgroup of $\mathrm{d} \Pi F$. The functor $\mathrm{dProd} 2 \operatorname{Prod}(F)$ yielding a homomorphism from $\mathrm{d} \Pi F$ to $\Pi \cup F$ is defined by
(Def. 10) for every element $x$ of $\mathrm{d} \Pi F$ and for every element $i$ of $I, x(i)=i t(x) \upharpoonright J(i)$.
Now we state the proposition:
(25) Let us consider a non empty set $I$, a non-empty, disjoint valued many sorted set $J$ indexed by $I$, a $J$-indexed family of group families $F$, an element $y$ of $\Pi \bigcup F$, and an element $i$ of $I$. Then $y \upharpoonright J(i) \in \Pi(F(i))$.
Proof: Set $x=y \upharpoonright J(i)$. Set $z=$ the support of $F(i)$. For every object $j$ such that $j \in J(i)$ holds $x(j) \in z(j)$ by [7, (49), (1)].
Let $I$ be a non empty set, $J$ be a non-empty, disjoint valued many sorted set indexed by $I$, and $F$ be a $J$-indexed family of group families. Note that $\mathrm{dProd} 2 \operatorname{Prod}(F)$ is bijective.

The functor $\operatorname{Prod} 2 \mathrm{dProd}(F)$ yielding a homomorphism from $\Pi \cup F$ to $\mathrm{d} \Pi F$ is defined by the term
(Def. 11) $(\operatorname{dProd} 2 \operatorname{Prod}(F))^{-1}$.
Now we state the proposition:
(26) Let us consider a non empty set $I$, a non-empty, disjoint valued many sorted set $J$ indexed by $I$, a $J$-indexed family of group families $F$, an element $x$ of $\Pi \cup F$, and an element $i$ of $I$. Then $x \upharpoonright J(i)=(\operatorname{Prod} 2 \operatorname{dProd}(F))(x)(i)$.
Let $I$ be a non empty set, $J$ be a non-empty, disjoint valued many sorted set indexed by $I$, and $F$ be a $J$-indexed family of group families. Note that $\operatorname{Prod} 2 \operatorname{dProd}(F)$ is bijective.
(27) Let us consider a non empty set $I$, a non-empty, disjoint valued many sorted set $J$ indexed by $I$, and a $J$-indexed family of group families $F$. Then $\operatorname{Prod} 2 \mathrm{dProd}(F)=(\mathrm{dProd} 2 \operatorname{Prod}(F))^{-1}$.
Let $I$ be a non empty set, $J$ be a non-empty, disjoint valued many sorted set indexed by $I, F$ be a $J$-indexed family of group families, and $x$ be a function. The functor rsupport $(x, F)$ yielding a disjoint valued many sorted set indexed by $I$ is defined by
(Def. 12) for every element $i$ of $I$, it $(i)=\operatorname{support}(x \upharpoonright J(i), F(i))$.
Now we state the propositions:
(28) Let us consider a non empty set $I$, a non-empty, disjoint valued many sorted set $J$ indexed by $I$, a $J$-indexed family of group families $F$, and a function $x$. Then support $(x, \bigcup F)=\bigcup \operatorname{rsupport}(x, F)$.
Proof: Set $y=\operatorname{rsupport}(x, F)$. For every object $j, j \in \operatorname{support}(x, \bigcup F)$ iff $j \in \bigcup y$ by (21), [7, (49), (3)], [9, (74)].
(29) Let us consider a non empty set $I$, a non-empty, disjoint valued many sorted set $J$ indexed by $I$, a $J$-indexed family of group families $F$, and functions $x, y, z$. Suppose $z \in \mathrm{~d} \prod F$ and $x=(\mathrm{dProd} 2 \operatorname{Prod}(F))(z)$. Then
(i) $\operatorname{rsupport}(x, F) \upharpoonright \operatorname{support}(z, \operatorname{SumBundle}(F))$ is a non-empty, disjoint valued many sorted set indexed by support ( $z, \operatorname{SumBundle}(F)$ ), and
(ii) $\operatorname{support}(x, \bigcup F)=\bigcup(\operatorname{rsupport}(x, F) \upharpoonright \operatorname{support}(z, \operatorname{SumBundle}(F)))$.

Proof: Set $s_{1}=\operatorname{rsupport}(x, F)$. Set $s_{2}=\operatorname{support}(z, \operatorname{SumBundle}(F))$. Set $f=s_{1} \upharpoonright s_{2}$. For every objects $s, t$ such that $s \neq t$ holds $f(s)$ misses $f(t)$ by [7, (47)]. $\emptyset \notin \operatorname{rng} f$ by [7, (47)], [10, (5)], [16, (44)]. support $(x, \cup F)=\bigcup s_{1}$. For every object $k$ such that $k \in \operatorname{support}(x, \bigcup F)$ holds $k \in \bigcup\left(s_{1} \upharpoonright s_{2}\right)$ by [10, (6)], [16, (44)], [18, (57)], [7, (47), (3)].
(30) Let us consider a non empty set $I$, a non-empty, disjoint valued many sorted set $J$ indexed by $I$, a $J$-indexed family of group families $F$, and a function $y$. Suppose $y \in \operatorname{sum} \bigcup F$. Then there exists a function $x$ such that
(i) $y=(\mathrm{dProd} 2 \operatorname{Prod}(F))(x)$, and
(ii) $x \in \mathrm{~d} \sum F$.

Proof: Consider $x$ being an element of $\Omega_{\mathrm{d}} \prod_{F}$ such that $y=(\mathrm{dProd} 2$ Prod $(F))(x)$. Set $s_{1}=\operatorname{rsupport}(y, F)$. $\operatorname{support}(y, \bigcup F)=\bigcup s_{1}$. For every element $i$ of $I, x(i) \in(\operatorname{SumBundle}(F))(i)$ by [7, (3)], [9, (74)], [12, (8)]. Set $S=\operatorname{SumBundle}(F)$. Reconsider $W=$ the support of $S$ as a many sorted set indexed by $I$. For every object $i$ such that $i \in I$ holds $x(i) \in W(i)$. Reconsider $s_{2}=s_{1} \upharpoonright \operatorname{support}(x, \operatorname{SumBundle}(F))$ as a non-empty, disjoint valued many sorted set indexed by support $(x, \operatorname{SumBundle}(F))$. $\bigcup s_{2}$ is finite. $\operatorname{dom} s_{2}$ is finite.
(31) Let us consider a non empty set $I$, a non-empty, disjoint valued many sorted set $J$ indexed by $I$, a $J$-indexed family of group families $F$, and functions $x, y$. Suppose $x, x \in \mathrm{~d} \sum F$. Then $(\mathrm{dProd} 2 \operatorname{Prod}(F))(x) \in \operatorname{sum} \bigcup F$. Proof: Reconsider $y=(\mathrm{dProd} 2 \operatorname{Prod}(F))(x)$ as an element of $\Pi \cup F$. Set $s_{1}=\operatorname{rsupport}(y, F)$. Reconsider $s_{2}=s_{1} \upharpoonright \operatorname{support}(x, \operatorname{SumBundle}(F))$ as a non-empty, disjoint valued many sorted set indexed by support $(x$, SumBundle $(F))$. For every object $i$ such that $i \in \operatorname{dom} s_{2}$ holds $s_{2}(i)$ is finite by [16, (40)], [7, (49)]. support $(y, \bigcup F)$ is finite.
(32) Let us consider a non empty set $I$, a non-empty, disjoint valued many sorted set $J$ indexed by $I$, and a $J$-indexed family of group families $F$. Then $\operatorname{rng}\left(\mathrm{dProd} 2 \operatorname{Prod}(F) \upharpoonright \mathrm{d} \sum F\right)=\Omega_{\text {sum }} \bigcup F$.
Proof: For every object $y, y \in \operatorname{rng}\left(\mathrm{dProd} 2 \operatorname{Prod}(F) \upharpoonright \Omega_{\mathrm{d} \sum F}\right)$ iff $y \in$ $\Omega_{\text {sum }} \bigcup_{F}$ by [18, (61)], (31), [7, (47)], (30).
Let $I$ be a non empty set, $J$ be a non-empty, disjoint valued many sorted set indexed by $I$, and $F$ be a $J$-indexed family of group families. The functor dSum2Sum $(F)$ yielding a homomorphism from $\mathrm{d} \sum F$ to sum $\bigcup F$ is defined by the term
(Def. 13) $\quad \mathrm{dProd} 2 \operatorname{Prod}(F) \upharpoonright \mathrm{d} \sum F$.
One can verify that dSum2Sum $(F)$ is bijective.

The functor Sum2dSum $(F)$ yielding a homomorphism from sum $\bigcup F$ to d $\sum F$ is defined by the term
(Def. 14) $(\text { dSum2Sum }(F))^{-1}$.
Now we state the proposition:
(33) Let us consider a non empty set $I$, a non-empty, disjoint valued many sorted set $J$ indexed by $I$, and a $J$-indexed family of group families $F$. Then $\operatorname{Sum} 2 \mathrm{dSum}(F)=\operatorname{Prod} 2 \mathrm{dProd}(F) \upharpoonright \operatorname{sum} \bigcup F$. The theorem is a consequence of (2).
Let $I$ be a non empty set, $J$ be a non-empty, disjoint valued many sorted set indexed by $I$, and $F$ be a $J$-indexed family of group families. One can check that $\operatorname{Sum} 2 \mathrm{dSum}(F)$ is bijective.

Now we state the proposition:
(34) Let us consider a non empty set $I$, a non-empty, disjoint valued many sorted set $J$ indexed by $I$, and a $J$-indexed family of group families $F$. Then dSum2Sum $(F)=(\operatorname{Sum} 2 d \operatorname{Sum}(F))^{-1}$.
Let $I$ be a non empty set, $G$ be a group, and $F$ be an internal direct sum components of $G$ and $I$. The functor $\operatorname{InterHom}(F)$ yielding a homomorphism from $\operatorname{sum} F$ to $G$ is defined by
(Def. 15) it is bijective and for every finite-support function $x$ from $I$ into $G$ such that $x \in \operatorname{sum} F$ holds $i t(x)=\prod x$.
Let $J$ be a non-empty, disjoint valued many sorted set indexed by $I, M$ be a direct sum components of $G$ and $I, N$ be a $J$-indexed family of group families, and $h$ be a many sorted set indexed by $I$. Assume for every element $i$ of $I$, there exists a homomorphism $h_{0}$ from (SumBundle $\left.(N)\right)(i)$ to $M(i)$ such that $h_{0}=h(i)$ and $h_{0}$ is bijective. The functor $\operatorname{ProdHom}(G, M, N, h)$ yielding a homomorphism from $\mathrm{d} \sum N$ to sum $M$ is defined by
(Def. 16) $\quad$ it $=\operatorname{SumMap}(\operatorname{SumBundle}(N), M, h)$ and it is bijective.
Now we state the propositions:
(35) Let us consider a non empty set $I$, a non-empty, disjoint valued many sorted set $J$ indexed by $I$, a group $G$, a direct sum components $M$ of $G$ and $I$, and a $J$-indexed family of group families $N$. Suppose for every element $i$ of $I, N(i)$ is a direct sum components of $M(i)$ and $J(i)$. Then $\bigcup N$ is a direct sum components of $G$ and $\bigcup J$.
Proof: Consider $f_{2}$ being a homomorphism from sum $M$ to $G$ such that $f_{2}$ is bijective. Define $\mathcal{P}$ (object) $=\Omega_{\text {sum }(N(\$ 1(\in I)))}$. Consider $D_{2}$ being a function such that dom $D_{2}=I$ and for every object $i$ such that $i \in I$ holds $D_{2}(i)=\mathcal{P}(i)$ from [7, Sch. 3]. Define $\mathcal{Q}$ (object) $=\Omega_{M\left(\$_{1}(\in I)\right)}$. Consider $R_{1}$ being a function such that dom $R_{1}=I$ and for every object $i$ such
that $i \in I$ holds $R_{1}(i)=\mathcal{Q}(i)$ from [7, Sch. 3]. Define $\mathcal{R}[$ object, object] $\equiv$ there exists a homomorphism $f_{3}$ from $\operatorname{sum}\left(N\left(\$_{1}(\in I)\right)\right)$ to $M\left(\$_{1}(\in I)\right)$ such that $f_{3}=\$_{2}$ and $f_{3}$ is bijective. For every element $i$ of $I$, there exists an element $y$ of $\bigcup D_{2} \dot{\rightarrow} \bigcup R_{1}$ such that $\mathcal{R}[i, y]$ by [7, (3)], [9, (74)]. Consider $f_{1}$ being a function from $I$ into $\bigcup D_{2} \dot{\rightarrow} \bigcup R_{1}$ such that for every element $i$ of $I, \mathcal{R}\left[i, f_{1}(i)\right]$ from [8, Sch. 3]. For every element $i$ of $I$, there exists a homomorphism $h_{0}$ from (SumBundle $\left.(N)\right)(i)$ to $M(i)$ such that $h_{0}=f_{1}(i)$ and $h_{0}$ is bijective.
(36) Let us consider a non empty set $I$, a non-empty, disjoint valued many sorted set $J$ indexed by $I$, a group $G$, an internal direct sum components $M$ of $G$ and $I$, and a $J$-indexed family of group families $N$. Suppose for every element $i$ of $I, N(i)$ is an internal direct sum components of $M(i)$ and $J(i)$. Then $\bigcup N$ is an internal direct sum components of $G$ and $\cup J$. Proof: Consider $f_{3}$ being a homomorphism from sum $M$ to $G$ such that $f_{3}$ is bijective and for every finite-support function $x$ from $I$ into $G$ such that $x \in \operatorname{sum} M$ holds $f_{3}(x)=\prod x$. Define $\mathcal{Q}[$ object, object $] \equiv$ there exists an internal direct sum components $N_{1}$ of $M\left(\$_{1}(\in I)\right)$ and $J\left(\$_{1}(\in I)\right)$ such that $N_{1}=N\left(\$_{1}\right)$ and $\$_{2}=\operatorname{InterHom}\left(N_{1}\right)$. For every object $x$ such that $x \in I$ there exists an object $y$ such that $\mathcal{Q}[x, y]$. Consider $f_{1}$ being a function such that dom $f_{1}=I$ and for every object $i$ such that $i \in I$ holds $\mathcal{Q}\left[i, f_{1}(i)\right]$ from [7, Sch. 2]. Set $f_{2}=\operatorname{ProdHom}\left(G, M, N, f_{1}\right)$. For every element $i$ of $I$, there exists a homomorphism $h_{0}$ from (SumBundle $\left.(N)\right)(i)$ to $M(i)$ such that $h_{0}=f_{1}(i)$ and $h_{0}$ is bijective and for every finite-support function $x$ from $J(i)$ into $M(i)$ such that $x \in(\operatorname{SumBundle}(N))(i)$ holds $h_{0}(x)=\prod x$. For every element $i$ of $I$, there exists a homomorphism $h_{0}$ from (SumBundle $(N))(i)$ to $M(i)$ such that $h_{0}=f_{1}(i)$ and $h_{0}$ is bijective. Reconsider $h=f_{3} \cdot f_{2} \cdot \operatorname{Sum} 2 d \operatorname{Sum}(N)$ as a homomorphism from sum $\cup N$ to $G$. Reconsider $U_{2}=\bigcup J$ as a non empty set. Reconsider $U_{3}=\bigcup N$ as a direct sum components of $G$ and $U_{2}$. For every object $j$ such that $j \in U_{2}$ holds $U_{3}(j)$ is a subgroup of $G$ by (21), [16, (56)]. For every finite-support function $x$ from $U_{2}$ into $G$ such that $x \in \operatorname{sum} U_{3}$ holds $h(x)=\prod x$ by [16, (42), (40)], [7, (13)], [8, (5), (15)].

## 4. Conservation Rule of Direct Sum Decomposition for Layering

Now we state the propositions:
(37) Let us consider a non empty set $I$, a non-empty, disjoint valued many sorted set $J$ indexed by $I$, a group $G$, a group family $M$ of $I$, and a $J$-indexed family of group families $N$. Suppose $\bigcup N$ is a direct sum components of $G$ and $\bigcup J$ and for every element $i$ of $I, N(i)$ is a direct sum
components of $M(i)$ and $J(i)$. Then $M$ is a direct sum components of $G$ and $I$.
Proof: Set $U_{3}=\bigcup N$. Consider $f_{4}$ being a homomorphism from sum $U_{3}$ to $G$ such that $f_{4}$ is bijective. Define $\mathcal{P}$ (object) $)=$ the carrier of $\operatorname{sum}\left(N\left(\$_{1}(\in\right.\right.$ $I))$ ). Consider $D_{2}$ being a function such that $\operatorname{dom} D_{2}=I$ and for every object $i$ such that $i \in I$ holds $D_{2}(i)=\mathcal{P}(i)$ from [7, Sch. 3]. Define $\mathcal{Q}($ object $)=$ the carrier of $M\left(\$_{1}(\in I)\right)$. Consider $R_{1}$ being a function such that $\operatorname{dom} R_{1}=I$ and for every object $i$ such that $i \in I$ holds $R_{1}(i)=\mathcal{Q}(i)$ from [7, Sch. 3]. Define $\mathcal{R}$ [object, object] $\equiv$ there exists a homomorphism $f_{3}$ from $M\left(\$_{1}(\in I)\right)$ to $\operatorname{sum}\left(N\left(\$_{1}(\in I)\right)\right)$ such that $f_{3}=\$_{2}$ and $f_{3}$ is bijective. For every element $i$ of $I$, there exists an element $y$ of $\cup R_{1} \dot{\rightarrow} \cup D_{2}$ such that $\mathcal{R}[i, y]$ by [17, (62), (63)], [7, (3)], [9, (74)]. Consider $f_{1}$ being a function from $I$ into $\bigcup R_{1} \dot{\rightarrow} \bigcup D_{2}$ such that for every element $i$ of $I$, $\mathcal{R}\left[i, f_{1}(i)\right]$ from [8, Sch. 3]. For every element $i$ of $I$, there exists a homomorphism $h_{0}$ from $M(i)$ to (SumBundle $\left.(N)\right)(i)$ such that $h_{0}=f_{1}(i)$ and $h_{0}$ is bijective.
(38) Let us consider a non empty set $I$, a non-empty, disjoint valued many sorted set $J$ indexed by $I$, a group $G$, a subgroup family $M$ of $I$ and $G$, and a $J$-indexed family of group families $N$. Suppose $\cup N$ is an internal direct sum components of $G$ and $\cup J$ and for every element $i$ of $I, N(i)$ is an internal direct sum components of $M(i)$ and $J(i)$. Then $M$ is an internal direct sum components of $G$ and $I$.
Proof: Reconsider $U_{2}=\bigcup J$ as a non empty set. Consider $f_{4}$ being a homomorphism from sum $\cup N$ to $G$ such that $f_{4}$ is bijective and for every finite-support function $x$ from $U_{2}$ into $G$ such that $x \in \operatorname{sum} \cup N$ holds $f_{4}(x)=\Pi x$. Define $\mathcal{Q}[$ object, object $] \equiv$ there exists an internal direct sum components $N_{1}$ of $M\left(\$_{1}(\in I)\right)$ and $J\left(\$_{1}(\in I)\right)$ such that $N_{1}=N\left(\$_{1}\right)$ and $\$_{2}=\left(\operatorname{InterHom}\left(N_{1}\right)\right)^{-1}$. For every object $x$ such that $x \in I$ there exists an object $y$ such that $\mathcal{Q}[x, y]$.

Consider $f_{1}$ being a function such that $\operatorname{dom} f_{1}=I$ and for every object $i$ such that $i \in I$ holds $\mathcal{Q}\left[i, f_{1}(i)\right]$ from [7, Sch. 2]. Reconsider $f_{3}=\operatorname{SumMap}\left(M,(\operatorname{SumBundle}(N)), f_{1}\right)$ as a homomorphism from sum $M$ to $\mathrm{d} \sum N$. For every element $i$ of $I$, there exists a homomorphism $h_{0}$ from $M(i)$ to $(\operatorname{SumBundle}(N))(i)$ such that $h_{0}=f_{1}(i)$ and $h_{0}$ is bijective by [17, (62), (63)]. Reconsider $h=f_{4} \cdot \mathrm{dSum} 2 \operatorname{Sum}(N) \cdot f_{3}$ as a homomorphism from sum $M$ to $G$. For every element $i$ of $I$, there exists a homomorphism $h_{0}$ from $(\operatorname{SumBundle}(N))(i)$ to $M(i)$ such that $h_{0}^{-1}=f_{1}(i)$ and $h_{0}$ is bijective and for every finite-support function $x$ from $J(i)$ into $M(i)$ such that $x \in(\operatorname{SumBundle}(N))(i)$ holds $h_{0}(x)=\prod x$. For every element $i$ of $I$, there exists a homomorphism $h_{0}$ from (SumBundle $\left.(N)\right)(i)$ to $M(i)$ such
that $h_{0}{ }^{-1}=f_{1}(i)$ and $h_{0}$ is bijective. For every finite-support function $x$ from $I$ into $G$ such that $x \in \operatorname{sum} M$ holds $h(x)=\prod x$ by [16, (40)], [7, (13)], [8, (5), (15)].
(39) Let us consider a non empty set $I_{2}$, and a group family $F_{2}$ of $I_{2}$. Suppose for every element $i$ of $I_{2}, \overline{\overline{F_{2}(i)}}=1$. Then $\overline{\bar{\alpha}}=1$, where $\alpha$ is the carrier of sum $F_{2}$.
Proof: For every object $x$ such that $x \in \Omega_{\text {sum } F_{2}}$ holds $x=\mathbf{1}_{\text {sum } F_{2}}$ by [16, (42)], [1, (30)], [2, (102)], [10, (5)].
(40) Let us consider a non empty set $I$, a group $G$, and a finite-support function $x$ from $I$ into $G$. Suppose for every object $i$ such that $i \in I$ holds $x(i)=\mathbf{1}_{G}$. Then $\prod x=\mathbf{1}_{G}$.
(41) Let us consider a non empty set $I$, a group $G$, a finite-support function $x$ from $I$ into $G$, and an element $a$ of $G$. If $I=\{1,2\}$ and $x=\left\langle a, \mathbf{1}_{G}\right\rangle$, then $\prod x=a$.
Proof: Reconsider $i_{1}=1$ as an element of $I$. Set $y=\left(I \longmapsto \mathbf{1}_{G}\right)+\cdot\left(i_{1}, a\right)$. For every object $i$ such that $i \in \operatorname{dom} x$ holds $x(i)=y(i)$ by [3, (44)], 4, (31), (32)], [15, (7)].
(42) Let us consider a group $G$, non empty sets $I_{1}, I_{2}$, a direct sum components $F_{1}$ of $G$ and $I_{1}$, and a group family $F_{2}$ of $I_{2}$. Suppose $I_{1}$ misses $I_{2}$ and for every element $i$ of $I_{2}, \overline{\overline{F_{2}(i)}}=1$. Then $F_{1}+\cdot F_{2}$ is a direct sum components of $G$ and $I_{1} \cup I_{2}$.
Proof: Reconsider $I=\{1,2\}$ as a non empty set. Set $J=\left\{\left\langle 1, I_{1}\right\rangle,\langle 2\right.$, $\left.\left.I_{2}\right\rangle\right\}$. For every objects $x, y_{1}, y_{2}$ such that $\left\langle x, y_{1}\right\rangle,\left\langle x, y_{2}\right\rangle \in J$ holds $y_{1}=y_{2} . \emptyset \notin \operatorname{rng} J$. For every objects $i, j$ such that $i \neq j$ holds $J(i)$ misses $J(j)$. Reconsider $M=\left\langle\operatorname{sum} F_{1}\right.$, sum $\left.F_{2}\right\rangle$ as a group family of $I$. $\overline{\overline{\Omega_{\text {sum } F_{2}}}}=1$. Consider $w$ being an object such that $\{w\}=\Omega_{\text {sum } F_{2}}$. For every functions $x, y$ such that $x, y \in \Omega \prod_{M}$ and $x(1)=y(1)$ holds $x=y$ by [12, (5)], [3, (44)].

Consider $h_{1}$ being a homomorphism from sum $F_{1}$ to $G$ such that $h_{1}$ is bijective. Set $C_{1}=$ the carrier of $\Pi M$. Set $C_{2}=$ the carrier of $G$. Define $\mathcal{P}$ [element of $C_{1}$, element of $\left.C_{2}\right] \equiv \$_{2}=h_{1}\left(\$_{1}(1)\right)$. For every element $x$ of $C_{1}$, there exists an element $y$ of $C_{2}$ such that $\mathcal{P}[x, y]$ by [12, (5)], [3, (44)], [8, (5)]. Consider $h$ being a function from $C_{1}$ into $C_{2}$ such that for every element $x$ of $C_{1}, \mathcal{P}[x, h(x)]$ from [8, Sch. 3]. For every objects $x_{1}, x_{2}$ such that $x_{1}, x_{2} \in C_{1}$ and $h\left(x_{1}\right)=h\left(x_{2}\right)$ holds $x_{1}=x_{2}$ by [12, (5)], [3, (44)], [8, (19)]. For every object $y$ such that $y \in C_{2}$ there exists an object $x$ such that $x \in C_{1}$ and $y=h(x)$ by [8, (11)], [3, (44)]. For every elements $a, b$ of $C_{1}, h(a \cdot b)=h(a) \cdot h(b)$ by [3, (44)], [12, (5)], [10, (1)]. Reconsider $M=\left\langle\operatorname{sum} F_{1}\right.$, sum $\left.F_{2}\right\rangle$ as a direct sum components
of $G$ and $I$. Set $N=\left\langle F_{1}, F_{2}\right\rangle$. For every element $i$ of $I, N(i)$ is a group family of $J(i)$ by [3, (44)]. For every element $i$ of $I, N(i)$ is a direct sum components of $M(i)$ and $J(i)$ by [3, (44)]. For every object $x$ such that $x \in \operatorname{dom} F_{1} \cup \operatorname{dom} F_{2}$ holds if $x \in \operatorname{dom} F_{2}$, then $(\cup N)(x)=F_{2}(x)$ and if $x \notin \operatorname{dom} F_{2}$, then $(\cup N)(x)=F_{1}(x)$ by (21), [3, (44)].
(43) Let us consider a group $G$, non empty sets $I_{1}, I_{2}$, an internal direct sum components $F_{1}$ of $G$ and $I_{1}$, and a subgroup family $F_{2}$ of $I_{2}$ and $G$. Suppose $I_{1}$ misses $I_{2}$ and $F_{2}=I_{2} \longmapsto\{\mathbf{1}\}_{G}$. Then $F_{1}+\cdot F_{2}$ is an internal direct sum components of $G$ and $I_{1} \cup I_{2}$.
Proof: Reconsider $I=\{1,2\}$ as a non empty set. Set $J=\left\{\left\langle 1, I_{1}\right\rangle,\langle 2\right.$, $\left.\left.I_{2}\right\rangle\right\}$. For every objects $x, y_{1}, y_{2}$ such that $\left\langle x, y_{1}\right\rangle,\left\langle x, y_{2}\right\rangle \in J$ holds $y_{1}=y_{2} . \emptyset \notin \operatorname{rng} J$. For every objects $i, j$ such that $i \neq j$ holds $J(i)$ misses $J(j)$. Reconsider $M=\left\langle G,\{\mathbf{1}\}_{G}\right\rangle$ as a group family of $I$. For every functions $x, y$ such that $x, y \in \Omega \Pi_{M}$ and $x(1)=y(1)$ holds $x=y$ by [12, (5)], [3, (44)]. Set $h_{1}=\operatorname{id}_{(\text {the carrier of } G)}$. Set $C_{1}=$ the carrier of $\Pi M$. Set $C_{2}=$ the carrier of $G$. Define $\mathcal{P}$ [element of $C_{1}$, element of $\left.C_{2}\right] \equiv$ $\$_{2}=h_{1}\left(\$_{1}(1)\right)$. For every element $x$ of $C_{1}$, there exists an element $y$ of $C_{2}$ such that $\mathcal{P}[x, y]$ by [12, (5)], [3, (44)], [8, (5)]. Consider $h$ being a function from $C_{1}$ into $C_{2}$ such that for every element $x$ of $C_{1}, \mathcal{P}[x, h(x)]$ from [ 8 , Sch. 3]. For every objects $x_{1}, x_{2}$ such that $x_{1}, x_{2} \in C_{1}$ and $h\left(x_{1}\right)=h\left(x_{2}\right)$ holds $x_{1}=x_{2}$ by [12, (5)], 3, (44)], [8, (19)]. For every object $y$ such that $y \in C_{2}$ there exists an object $x$ such that $x \in C_{1}$ and $y=h(x)$ by [8, (11)], [3, (44)]. For every elements $a, b$ of $C_{1}, h(a \cdot b)=h(a) \cdot h(b)$ by [3, (44)], [12, (5)], [10, (1)].

Reconsider $M=\left\langle G,\{\mathbf{1}\}_{G}\right\rangle$ as a direct sum components of $G$ and $I$. For every element $i$ of $I, M(i)$ is a subgroup of $G$ by [3, (44)], [16, (54)]. For every finite-support function $x$ from $I$ into $G$ such that $x \in \operatorname{sum} M$ holds $h(x)=\Pi x$ by [10, (9)], [3, (44)], (41). Set $N=\left\langle F_{1}, F_{2}\right\rangle$. For every element $i$ of $I, N(i)$ is a group family of $J(i)$ by [3, (44)]. For every element $i$ of $I$, $N(i)$ is an internal direct sum components of $M(i)$ and $J(i)$ by [3, (44)], [15, (7)], [1, (30)], (39). For every object $x$ such that $x \in \operatorname{dom} F_{1} \cup \operatorname{dom} F_{2}$ holds if $x \in \operatorname{dom} F_{2}$, then $(\cup N)(x)=F_{2}(x)$ and if $x \notin \operatorname{dom} F_{2}$, then $(\cup N)(x)=F_{1}(x)$ by (21), [3, (44)].

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