# A THEOREM CONCERNING A RESTRICTED RULE OF SUBSTITUTION IN THE FIELD OF PROPOSITIONAL CALCULI. I 

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In this paper the rule of simultaneous substitution ordinarily used in the field of propositional calculi will be called restricted if, in the formalization of the given system, its applications are limited to the axioms of that system. So far as I know, A. Lindenbaum was the first who investigated an instance of this rule. Namely, around 1934 he informed me casually that there are some systems whose axiomatizations have a special structure of the bi-valued propositional calculus in which a replacement of the rule of simultaneous substitution by the restricted one does not affect the strength of these systems. Since Lindenbaum never published his research concerning this and related results, I have no idea exactly how his theorem was formulated and how it was proved. Much later, in [1], pp. 148-151, section 27 (see especially p. 150), A. Church sketches a proof of a theorem which states that any system of the classical propositional calculus or any partial system of that calculus whose only rules of procedure are: detachment for implication and substitution (not necessarily simultaneous) may be reformulated into a system which has the same theorem as the original one and whose single rule of procedure is detachment. An inspection of Church's proof of this theorem shows that it holds simply through replacing each axiom of a system under consideration by the corresponding axiom schema. Since, certainly, Lindenbaum did not intend to reject the rule of substitution totally in formulating his theorem and, probably, he did not use the axiom schemata in the deductions which were needed for a proof of the theorem, the theorems discussed above are rather distinctly different.

In this note we will prove the following theorem concerning the restricted rule of simultaneous substitution:

Theorem A If (i) $T$ is an arbitrary, consistent propositional system whose formalization satisfies the conditions:
(a) The set of primitive notions of $T$ contains at least the proposition
forming functor for two propositional arguments $C$ which does not necessarily coincide with the classical implication;
(b) The set of the rules of procedure of $T$ contains at least the following two rules:

R1 The rule of simultaneous substitution which is ordinarily used in the field of propositional calculi, but here is adjusted to the primitive notions of system $T$,
and
R2 The rule of detachment in regard to functor C: If the formulas $\alpha$ and $C \alpha \beta$ are the theses of $T$, then formula $\beta$ can be added to $T$, as its new thesis;
(c) System $T$ is axiomatizable and its complete axiom set $\mathfrak{A}$, which can be finite or infinite, does not contain the axiom schemata,
(ii) Sequence $\mathbf{A}=\left\{\mathbf{a}_{1}, \ldots, \mathbf{a}_{m}\right\}, 1 \leqslant m<\infty$, is an arbitrary, unempty, finite subset which can be proper or unproper of the axiom set $\mathfrak{M}$,
(iii) Formula $\mathbf{b}$ which is a well-formed formula in the field of system $T$ and is not equiform with any term of sequence $\mathbf{A}$,
and
(iv) Rule R1* is the restricted rule RI,
then in T: Formula $\mathbf{b}$ is provable in the field of $\mathbf{A}$ by the applications of the rules R1 and R2, if and only if, formula $\mathbf{b}$ is provable in the field of $\mathbf{A}$ by the applications of the rules R1* and R2.

Obviously, Theorem A is a very strong generalization of Lindenbaum's theorem mentioned above. It should be noted that the axioms of $T$ are entirely undefined (clearly, we know only that at least one of the axioms belonging to sequence $A$ must have a form $C \alpha \beta$ ), and it is self-evident that Theorem A holds only for the rule of simultaneous substitution.

Proof:
1 In order to prove Theorem A in the most compact way I shall use, in the reasonings presented below, the following abbreviations:
(a) The abbreviations " $\alpha \approx \beta$," " $\left.A\right|_{R 1} \alpha$," and ' $\left.\{\alpha\}\right|_{R T} \beta$ ', will mean 'formula $\alpha$ is equiform with formula $\beta$," "in the field of sequence $A$ formula $\alpha$ is provable by $R I$ " and " $\beta$ is a consequence of $\alpha$ by RI" respectively. The analogous and obvious meanings will have the following abbreviations: "A $\left.\right|_{\overline{R 1 *}} \alpha$, " " $\left.A\right|_{R 1, R 2} \alpha$," " $\left.A\right|_{R 1^{*}, R 2} \alpha$," " $\left.\{\alpha\}\right|_{\overline{R 1 *}} \beta$," " $\left.\{\alpha, \beta\}\right|_{R 2} \gamma$," and " $\left.A\right|_{R 1^{*}} V$," and so forth. The last abbreviation given above means "in the field of $A$ the set (the sequence) of the formulas $V$ is provable by R1*,"
and the following tacit assumptions:
(b) In any sequence which will be considered below each of its terms occurs without repetition.
(c) If the sequence $\mathbf{Z}$ considered below contains three terms $\tau_{i}, \tau_{j}$, and $\tau_{f}$, $i<j<f$, such that in $\left.\mathbf{Z}\left\{\tau_{i}\right\}\right|_{\left.\right|_{\mathrm{R} 1}} \tau_{j}$ and $\left.\left\{\tau_{j}\right\}\right|_{\left.\right|_{R 1}} \tau_{f}$, then tacitly $\mathbf{Z}$ is substituted by a sequence which possesses exactly the same terms and ordering, but in which $\left.\left\{\tau_{i}\right\}\right|_{\overline{R 1}} \tau_{f}$. It is similar if instead of R1 we have RI*. Clearly, we can do it, since our rules of substitution are simultaneous.
(d) If a subsequence $V$ of the given sequence $\mathbf{Z}$ is such that it contains all such and only such terms of $\mathbf{Z}$ that satisfy all conditions of the given property $\phi$ and a subsequence $\mathbf{W}=\mathbf{Z}-\mathrm{V}$, then we assume tacitly that these two subsequences are ordered according to the order in which their terms occur in the sequence $\mathbf{Z}$. Moreover, we assume that in $\mathbf{Z}$ its terms are automatically rearranged in such a way that each term of V precedes every term of $\mathbf{W}$. Thus, in such a case we have: $\mathbf{Z}=\{\mathbf{V} ; \mathbf{W}\}$. Obviously, we can do this only under conditions where the subsequences $\mathbf{V}$ and $\mathbf{W}$ are disjoint and $\phi$ is suitably defined.
(e) If a subsequence $\mathbf{V}=\left\{\sigma_{1}, \ldots, \sigma_{m}\right\}, 0 \leqslant m<\infty$, of the given sequence $\mathbf{Z}$ is such that it contains all such and only such terms of $\mathbf{Z}$ which satisfy all conditions of the given property $\phi$, and formula $\tau$ also satisfies these conditions, but it is not a term of $\mathbf{Z}$, then the sequence $\mathbf{V}^{*}=\left\{\sigma_{1}, \ldots, \sigma_{m}\right.$, $\left.\sigma_{m+1}\right\}, 0 \leqslant m+1<\infty$, in which $\sigma_{m+1} \approx \tau$, is called an augmentation of V . And, if in $\mathbf{Z}$ we replace $\mathbf{V}$ by $V^{*}$, then this new sequence will be indicated by $\mathbf{Z *}$. Moreover, we tacitly assume that the augmentation of $\mathbf{Z}$ is always done in such a way that, if $\rho$ is the last term of $\mathbf{Z}$, then it is also the last term of $\mathbf{Z} *$. Thus, e.g., if $\mathbf{Z}=\{\mathbf{V} ; \mathbf{W}\}$, then $\mathbf{Z}^{*}=\{\mathbf{V} * ; \mathbf{W}\}$. It is self-evident that if, in the proof given below, it will be established that the given sequence $\mathbf{P}=\{\mathbf{Q} ; \mathbf{R}\}$ possesses a certain required property $\psi$ and its subsequence $\mathbf{Q}$ is augmented to $Q^{*}$ by a formula which does not contradict $\psi$, then $P^{*}=\left\{Q^{*} ; R\right\}$ also satisfies $\psi$.

2 Now, let us assume the antecedent of Theorem A. Since RI* is a restriction of R1, we know at once that it is sufficient to prove: In $T$, if $\left.\mathbf{A}\right|_{R 1, R 2} \mathbf{b}$, then $\left.\mathbf{A}\right|_{R 1^{*}, R 2} \mathbf{b}$. Hence, assume that in $\left.T \mathbf{A}\right|_{R 1, R 2} \mathbf{b}$. Then it follows from this assumption and the standard definition of a proof that there exists an unempty, finite sequence of the formulas:

$$
\mathfrak{(})=\left\{\mathbf{a}_{1}, \ldots \mathbf{a}_{m}, \mathbf{b}_{m+1}, \ldots, \mathbf{b}_{z}\right\}, 1<z<\infty
$$

constructed in accordance with the points (b) and (c) of 1 and such that:
(1) The terms $\boldsymbol{a}_{1}, \ldots, \boldsymbol{a}_{m}$ of $(\mathcal{P}$ are the axioms of $T$ which belong to A;
(2) The terms of subsequence $B=\left\{b_{m+1}, \ldots, b_{z}\right\}$ of $\mathcal{P}$ are such that if $\mathbf{b}_{t}, m+1 \leqslant t \leqslant z$, is a term of $\mathbf{B}$, then either
$(\alpha)$ there is a term $\sigma$ of $\mathfrak{P}$ which precedes $\mathbf{b}_{t}$ and $\left.\{\sigma\}\right|_{R 1} \mathbf{b}_{t}$,
( $\beta$ ) there are the terms $\sigma$ and $\tau$ of $\boldsymbol{\mathcal { D }}$ which precede $\mathbf{b}_{t}$ and $\left.\{\sigma, \tau\}\right|_{\mathrm{R}^{2}} \mathbf{b}_{t}$;
(3) The last term of $\mathfrak{Q}$, viz. $\mathrm{b}_{z}$, is such that $\mathrm{b}_{z} \approx \mathrm{~b}$.
2.1 Since it will be more convenient for our further reasoning, I
renumerate the terms of $B$, as follows: $B=\left\{\mathbf{B}_{m+1}, \ldots, \boldsymbol{b}_{z}\right\}=\left\{\mathbf{b}_{1}, \ldots, \mathbf{b}_{w}\right\}$, $1 \leqslant w<z, \mathbf{b}_{w} \approx \mathbf{b}$. Moreover, since it follows from points (ii) and (iii) of the antecedent of Theorem $A$ that both subsequences $A$ and $B$ of $\mathfrak{T}$ are unempty and since, according to the definition of $\mathfrak{P}$, they are disjoint, we have: $\mathfrak{T}=\{\mathbf{A} ; \mathbf{B}\}, c f ., 1$ (d).

3 The subsequences of B which will be defined below and which, eventually, can be empty, will be analyzed and used in our proof.
3.1 Let $\mathrm{V}_{1}=\left\{\alpha_{1}, \ldots, \alpha_{w}\right\}, 0 \leqslant n \leqslant w$, be a subsequence of B containing all such and only such terms of $\mathbf{B}$ that for every $\alpha_{j}, 1 \leqslant j \leqslant w, \alpha_{j}$ is a term of $\mathrm{V}_{1}$ if and only if $\alpha_{j}$ is a term of $\mathfrak{3}$ and in $\mathbf{A}$ there is a term $\boldsymbol{a}_{i}, 1 \leqslant i \leqslant m$, such that $\left.\left\{\mathbf{a}_{i}\right\}\right|_{R 1} \alpha_{j}$.
3.1.1 If $V_{1}$ is unempty, then since all terms of $V_{1}$ are generated by the application of R1 to the axioms of $T$ belonging to $\mathbf{A}$, it is self-evident that $\left.A\right|_{\mathrm{RI}^{1}} V_{1}$. Therefore, if $\mathrm{B}=\mathrm{V}_{1}$, then, since B is not empty, $c f$., 2.1, we have, obviously, $V_{1}=\{b\}$, i.e., that $\left.A\right|_{R^{1}} \times b$. And, in such a case Theorem $A$ is proved. On the other hand, if $\mathrm{B} \neq \mathrm{V}_{1}$, then for $\mathrm{C}=\mathrm{B}-\mathrm{V}_{1}$ we have, $c f_{\text {. }}, \mathbf{1}$ (c), $B=\left\{\mathbf{V}_{1} ; \mathbf{C}\right\}$, i.e., $\mathcal{P}=\left\{\mathbf{A} ; \mathbf{V}_{1} ; \mathbf{C}\right\}$, where $\mathbf{V}_{1}$ can be empty.
3.2 If $\mathbf{C}$ is unempty, let $\mathbf{V}_{2}=\left\{\beta_{1}, \ldots, \beta_{p}\right\}, 1 \leqslant p \leqslant w$, be a subsequence of $\mathbf{C}$ containing all such and only such elements of $C$ that for every $\beta_{k}, 1 \leqslant k \leqslant w$, $\beta_{k}$ is a term of $V_{2}$ if and only if $\beta_{k}$ is a term of $\mathcal{T}$ and in $\mathfrak{D}$ there are two terms $\sigma$ and $\tau$ which precede the first term of $C$ and such that $\tau \approx C \sigma \beta_{k}$ and $\left.\{\sigma, \tau\}\right|_{\bar{R} 2} \beta_{k}$.
3.2.1 If $\mathbf{V}_{2}=\mathbf{C}$, then $\mathcal{D}=\left\{\mathbf{A} ; \mathbf{V}_{1} ; \mathbf{V}_{2}\right\}$. Therefore, in such a case if $\mathbf{V}_{1}$ is empty, then $A \vdash_{\overline{R 2}} \mathbf{V}_{2}$, i.e., obviously, cf., 3.2, $A \vdash_{\overline{R 2}} \mathbf{b}$, and if $\mathrm{V}_{1}$ is unempty, then, cf., 3.1.1, $\left.\mathbf{A}\right|_{R I^{*}, R^{2}} \mathbf{V}_{2}$, i.e., clearly, $\left.\mathbf{A}\right|_{R 1^{*}, R_{2}} \mathbf{b}$. Thus, if $\mathfrak{T}=\left\{\mathbf{A} ; \mathbf{V}_{1} ; \mathbf{V}_{2}\right\}$, then Theorem $A$ is proved. On the other hand, if $V_{2} \neq C$, we have for $\mathbf{D}=\mathbf{C}-\mathbf{V}_{2}, c f ., 1(\mathrm{~d}), \mathrm{C}=\left\{\mathbf{V}_{2} ; \mathbf{D}\right\}$, i.e., $\mathfrak{D}=\left\{\mathbf{A} ; \mathbf{V}_{1} ; \mathbf{V}_{2} ; \mathbf{D}\right\}$. In such a case, since $D$ is unempty, its first term, $d_{1}$, neither
(a) can be such that $A \digamma_{R 1} d_{1}$, since otherwise it would be a term of $\mathrm{V}_{1}, c f$., 3.1, or for the same reason, $c f ., 1$ (c), such that $V_{1} \frac{1}{R 1} d_{1}$,
nor
(b) can be such that it would be a consequence by R2 of two terms belonging to $\boldsymbol{T}^{2}$ which precede the first term of $\mathrm{V}_{2}$, since otherwise it would be a term of $\mathrm{V}_{2}, c f ., 3.2$.

Hence, if $D$ is unempty, and $d_{1}$ is its first term, then either in $V_{2}$ there is a term $\sigma$ such that $\left.\{\sigma\}\right|_{R 1} d_{1}$ or in $\mathfrak{D}$ there are two terms $\sigma$ and $\tau$ such that they precede $d_{1}$, at least one of them is a term of $V_{2}, \tau \approx C \sigma d_{1}$ and $\left.\{\sigma, \tau\}\right|_{R_{2}} d_{1}$. Therefore, if in $\mathfrak{T}$ its subsequence $D$ is not empty, then also $\mathbf{V}_{2}$ is unempty.
3.3 If $\mathbf{D}$ is unempty, let $\mathbf{E}=\left\{\gamma_{1}, \ldots, \gamma_{q}\right\}, 1 \leqslant q<w$, be a subsequence of $\mathbf{D}$ containing all such and only such terms of $\mathbf{D}$ that for every $\gamma_{k}, 1 \leqslant k<w, \gamma_{k}$ is a term of $\mathbf{E}$ if and only if $\gamma_{k}$ is a term of $\mathcal{P}$ and in $\mathbf{V}_{2}$ there is a term $\sigma$ such that $\left.\{\sigma\}\right|_{\bar{R} 1} \gamma_{k}$.
3.4 If $D \neq E$, let $F=D-E$. Hence, if $E$ is unempty, $D=\{\mathbf{E} ; \mathbf{F}\}$, i.e., (2) $=\left\{\mathbf{A} ; \mathbf{V}_{1} ; \mathbf{V}_{2} ; E ; \mathbf{F}\right\}$. It follows at once from the definitions of $D$ and $E$ $c f$., 3.2.1, and 3.3, that if $F$ is not empty, then its first term, say $\mathbf{f}_{1}$, is such that in $\mathcal{T}$ there are two terms $\sigma$ and $\tau$ such that they precede $f_{1}, \tau \approx C \sigma f_{1}$ and $\left.\{\sigma, \tau\}\right|_{\bar{R} 2} f_{1}$. Moreover, clearly, cf., 3.2.1., at least one of these terms must be either a term of $V_{2}$ or a term of $E$, since otherwise $f_{1}$ would be a term of $\mathrm{V}_{2}$.

4 In this section it will be shown that if subsequence $E$ of $D$ is unempty, then we are always able to replace $\mathfrak{( 2 )}$ by its augmentation constructed effectively:

$$
\mathfrak{O}^{*}=\left\{\mathbf{A} ; \mathrm{V}_{1}^{*} ; \mathrm{V}_{2}^{*} ; \mathbf{F}\right\}
$$

in which its subsequences $A$ and $F$ are exactly the same as in $\mathfrak{P}$ and the subsequences $V_{1}^{*}$ and $V_{2}^{*}$ are such augmentations of $V_{1}$ and $V_{2}$ that each formula which in $\mathfrak{T}$ is a term of $\mathbf{E}$ occurs in $\mathfrak{D}^{*}$, as a term of $\mathbf{V}_{2}^{*}$.
4.1 Let us assume that in $\mathfrak{T}$ its subsequence $\mathbf{E}$ is not empty and, moreover, that $\gamma_{k}, 1 \leqslant k \leqslant q$, is an arbitrary term of $\mathbf{E}$. Then, according to the definition of $\mathbf{E}, c f$., 3.3, in $\mathbf{V}_{2}$ which is not empty, $c f$., 3.2.1, there is a term $\beta_{h}, 1 \leqslant h \leqslant p$, such that $\left.\left\{\beta_{h}\right\}\right|_{R 1} \gamma_{k}$. Since $\beta_{h}$ is a term of $\mathbf{V}_{2}$, in $\mathfrak{P}, c f$., 3.2, there are two such terms $\sigma$ and $\tau$ that precede the first term of $V_{2}$, such that $\tau \approx C \sigma \beta_{h}$, and $\left.\{\sigma, \tau\}\right|_{\overline{R 2}} \beta_{h}$. Hence, in accordance with the definition of $\mathrm{V}_{2}$, there are four possibilities: either both $\sigma$ and $\tau$ are the terms of A , or $\sigma$ is a term of $A$ and $\tau$ is a term of $V_{1}$, or $\sigma$ is a term of $V_{1}$ and $\tau$ is a term of $\mathbf{A}$, or both $\sigma$ and $\tau$ are the terms of $\mathrm{V}_{1}$. Consequently, we have to analyze the four possible cases:

Case 1. Both $\sigma$ and $\tau$ are the terms of $\mathrm{A}, \tau \approx C \sigma \beta_{h},\left.\{\sigma, \tau\}\right|_{\overline{\mathrm{R} 2}} \beta_{h}$ and $\left.\left\{\beta_{h}\right\}\right|_{R 1} \gamma_{k}$. Since both rules of substitution mentioned in the formulation of Theorem A are simultaneous, it follows at once from our present assumptions and the definition of $\mathrm{V}_{1}, c f ., 3.1$, and 3.1.1, that
(1) there must exist a formula $\mu$ such that $\left.\{\tau\}\right|_{\left.\right|_{\mathbb{R}}{ }^{*}} \mu$ and $\mu \approx C \rho \gamma_{k}$, and that
(2) either $\sigma \approx \rho$ or $\left.\sigma\right|_{R 1 *} \rho$.

Hence, we have to investigate the four obvious subcases:
Subcase 1a. $\sigma \approx \rho$ and $\mu$ is a term of $\mathfrak{T}$. Whence, if follows from point (1) and the definition of $V_{1}$ that $\mu$ is a term of $V_{1}$. However, such a case is impossible, since otherwise, cf., point (1) and the definition of $\mathrm{V}_{2}, \gamma_{k}$ would be a term of $V_{2}$.

Subcase $1 \mathrm{~b} . \sigma \approx \rho$ and $\mu$ is not a term of $\mathfrak{T}$. Hence, also $\mu$ is not a term of $\mathrm{V}_{1}$. Now, define:
(a) $\mathbf{V}_{1}^{* 1}=\left\{\mathbf{V}_{1} ; \mu\right\}=\left\{\alpha_{1}, \ldots, \alpha_{n}, \alpha_{n+1}\right\}, 1 \leqslant n+1, \alpha_{n+1} \approx \mu$.
(b) $\mathbf{V}_{2}^{* 1}=\left\{\mathbf{V}_{2} ; \gamma_{k}\right\}=\left\{\beta_{1}, \ldots, \beta_{p}, \beta_{p+1}\right\}, 1 \leqslant p+1, \beta_{p+1} \approx \gamma_{k}$.
(c) $\mathbf{E}^{0}=\left\{\gamma_{1}, \ldots, \gamma_{k-1}, \gamma_{k+1}, \ldots, \gamma_{q}\right\}, 1 \leqslant k \leqslant q$.

Now, we replace in $\mathfrak{B}$ its subsequences $V_{1}, V_{2}$, and $E$ by $V_{1}^{* 1}$, $V_{2}^{* 1}$, and $E^{0}$ respectively obtaining in such a way a new sequence $\mathfrak{D}^{* 1}=\left\{\mathbf{A} ; \mathbf{V}_{1}^{* 1} ; \mathbf{V}_{2}^{* 1}\right.$; $\left.\mathbf{E}^{0} ; \mathbf{F}\right\}$. Since, by assumptions, both $\sigma$ and $\tau$ are the terms of $\mathbf{A},\left.\{\tau\}\right|_{\overline{R \top}} \mu$, $\mu \approx C \rho \gamma_{k}$ and $\sigma \approx \rho$, we have $\mu \approx C \sigma_{k}$ and, therefore, $\left.\{\sigma, \mu\}\right|_{R^{2}} \gamma_{k}$. Hence, since in $\mathrm{V}_{1}^{* 1}$ its last term $\alpha_{n+1} \approx \mu \approx C \rho \gamma_{k} \approx C \sigma_{\gamma_{k}}$ and in $\mathrm{V}_{2}^{* 1}$ its last term $\beta_{p+1} \approx \gamma_{k}$, in $\left.\mathfrak{D}^{* 1} A\right|_{\overline{R 1 *}, R_{2}^{2}} \gamma_{k}$ while in $\left.\mathfrak{D} \approx A\right|_{\overline{R 1, R 2}} \gamma_{k}$. Thus, it is self-evident that if this subcase of Case 1 holds for $\gamma_{k}$, then we can always solve it accepting. instead of $\mathfrak{\bullet}$ its augmentation $\mathfrak{D}^{* 1}$ as a proof sequence of $b$.
Subcase 1c. $\left.\sigma\right|_{\mathrm{Rl}^{*}} \rho$ and in $\mathfrak{T} \mu$ is a term of $\mathrm{V}_{1}$. The case that $\rho$ is a term of $\mathfrak{T}$, i.e., $\rho$ is a term of $V_{1}$ is impossible, since otherwise $\gamma_{k}$ would be a term of $V_{2}$. Therefore, $\rho$ is not a term of $\mathfrak{D}$. Hence, define:
(d) $\mathrm{V}_{1}^{* 2}=\left\{\mathrm{V}_{1} ; \rho\right\}=\left\{\alpha_{1}, \ldots, \alpha_{n}, \alpha_{n+1}\right\}, 1 \leqslant n+1, \alpha_{n+1} \approx \rho$.

Then, using $\mathrm{V}^{*}$ and $\mathrm{E}^{0}$ as defined in points (b) and (c) above, we replace (2) by its augmentation $\mathfrak{D}^{* 2}=\left\{A ; V_{1}^{* 2} ; V_{2}^{* 1} ; E^{0} ; F\right\}$. Since $\tau$ is a term of $A$ and $\left.\{\tau\}\right|_{\mathbb{R}^{*}} \mu$, in $\left.\mathfrak{O}^{* 2} \mathbf{A}\right|_{\overline{R^{*}, R^{2}} \gamma_{k}}$. Hence, if this subcase of Case 1 holds for $\gamma_{k}$, then we can always solve it accepting instead of $\mathfrak{D}^{(2)}$ its augmentation $\mathfrak{D}^{* 2}$ as a proof sequence of $\mathbf{b}$.
Subcase 1d. $\left.\{\sigma\}\right|_{\mathbb{R T}^{*} \rho}$ and $\mu$ is not a term of © Whence, $\mu$ is also not a term of $V_{1}$. On the other hand, since either $\rho$ is a term of $\mathfrak{T}$ or $\rho$ is not a term of $\mathfrak{B}$, each of these two possibilities must be assumed and investigated separately. Wherefore:
Subcase $1 \mathrm{~d}_{1} .\left.\{\sigma\}\right|_{\overline{R 1}+} \rho ; \rho$ is a term of $\boldsymbol{P}$ and $\mu$ is not a term of $\mathbf{V}_{1}$. Since $\sigma$ is a term of $\mathbf{A},\left.\{\sigma\}\right|_{\overline{R I}_{1}} \rho$ and $\rho$ is a term of $\mathfrak{T}$, it follows from the definition of $\mathbf{V}_{1}, c f$., 3.1 and 3.1.1, that $\rho$ is a term of $\mathbf{V}_{1}$. Hence, define:
(e) $\mathbf{V}_{1}^{* 3}=\left\{\mathbf{V}_{1}, \mu\right\}=\left\{\alpha_{1}, \ldots, \alpha_{n}, \alpha_{n+1}\right\}, 1 \leqslant n+1, \alpha_{n+1} \approx \mu$.

Then, we replace $\mathfrak{D}^{(2)}$ by its augmentation $\mathfrak{D}^{* 3}=\left\{A ; V_{1}^{* 3} ; V_{2}^{* 1} ; E^{0} ; F\right\}$. Since both terms $\sigma$ and $\tau$ are the terms of $A,\left.\{\sigma\}\right|_{\overline{R 1}^{*}} \rho,\left.\{\tau\}\right|_{\bar{R} 1^{*}} \mu, \rho$ is a term of $\mathrm{V}_{1}$, i.e., clearly, $\rho$ is a term of $\mathrm{V}_{1}^{* 3}$, and, moreover, $\mu$ is a term of $\mathrm{V}_{1}^{* 3}$, in $\left.\mathfrak{D}^{* 3} \mathbf{A}\right|_{\mathbb{R 1}^{\star}, R_{2}} \gamma_{k}$. Therefore, if subcase $1 d_{1}$ holds for $\gamma_{k}$, then we can always solve it accepting instead of $\mathfrak{P}$ its augmentation $\mathfrak{Q}^{* 3}$ as a proof sequence of $b$.

Subcase $1 \mathrm{~d}_{2}$. $\{\sigma\}_{\overline{R 1}{ }^{\star}} \rho, \rho$ is not a term of $\mathcal{D}$ and $\mu$ is not a term of $\mathbf{V}_{1}$. Hence, $\rho$ is not a term of $\mathrm{V}_{1}$. Now, if in $\mathrm{A} \sigma$ precedes $\tau$, we define
(f) $\mathbf{V}_{1}^{* 4}=\left\{\mathbf{V}_{1}, \rho, \mu\right\}=\left\{\alpha_{1}, \ldots, \alpha_{n}, \alpha_{n+1}, \alpha_{n+2}\right\}, 1 \leqslant n+1, \alpha_{n+1} \approx \rho$ and $\alpha_{n+2} \approx \mu$.

On the other hand, if in $\mathbf{A} \tau$ precedes $\sigma$, then we define $\mathrm{V}_{1}^{* 4}$ as follows:
(g) $\mathbf{V}_{1}^{* 4}=\left\{\mathbf{V}_{1}, \mu, \rho\right\}=\left\{\alpha_{1}, \ldots, \alpha_{n}, \alpha_{n+1}, \alpha_{n+2}\right\}, 1 \leqslant n+1, \alpha_{n+1} \approx \mu$ and $\alpha_{n+2} \approx \rho$.

Remark I: Clearly, each of the sequences which are defined in points (f) and (g) above can be used in order to obtain a solution of subcase $1 \mathrm{~d}_{2}$. Since the obtained solution must be unique, it must be uniquely determined which of the sequences presented above should be accepted in regard to the term $\gamma_{k}$ under consideration. Since the assumptions concerning $\gamma_{k}$ inform
only that both $\sigma$ and $\tau$ are the terms of $\mathbf{A},\{\sigma\}_{\overline{R 1} \star \rho},\left.\{\tau\}\right|_{\left.\right|_{R 1}} \mu$, both $\rho$ and $\mu$ are not the terms of $\mathrm{V}_{1}$ and $\left.\{\rho, \mu\}\right|_{\overline{R 2}} \gamma_{k}$, the order of $\sigma$ and $\tau$ in $\mathbf{A}$ is only one possible determinent which uniquely selects a proper sequence in regard to $\gamma_{k}$.

Then, we replace $\mathfrak{( 2 )}$ by its augmentation $\mathfrak{D}^{* 4}=\left\{\mathbf{A} ; \mathbf{V}_{1}^{* 4} ; \mathbf{V}_{2}^{* 1} ; \mathbf{E}^{0} ; \mathbf{F}\right\}$ in which $\mathrm{V}_{1}^{* 4}$ is defined either as in (f) or as in (g) according to Remark I. Since both $\sigma$ and $\tau$ are terms of $\mathbf{A},\left.\{\sigma\}\right|_{\overline{R 1 *}} \rho$ and $\left.\{\tau\}\right|_{\bar{R} 1^{*}} \mu$, in $\mathfrak{D}^{* 4}$ both $\rho$ and $\mu$ are the terms of $\mathrm{V}_{1}^{* 4}$. Whence, in $\left.\mathfrak{D}^{* 4} A\right|_{\mathrm{RI}^{*}, \mathrm{R}_{2}} \gamma_{k}$. Therefore, if subcase $1 \mathrm{~d}_{2}$ holds for $\gamma_{k}$, then we can always solve it accepting instead of $\mathfrak{D}$ its augmentation $\mathfrak{D}^{* 4}$ as a proof sequence of $b$.

Thus, subcase 1 d is solved because it is established above that for each its possible instances, i.e., subcases $1 \mathrm{~d}_{1}$ and $1 \mathrm{~d}_{2}$, cf., also Remark I, we are able to construct in an effective way such unique augmentation of $\mathfrak{P}$, viz. $\mathfrak{D}_{\text {Sld }}^{*}$, that $\mathfrak{T}_{\text {Sld }}^{*}$ is a proof sequence of $b$ and that in $\left.\mathfrak{P}_{\text {Sld }}^{*} \mathbf{A}\right|_{\overline{R 1}{ }^{*}, R_{2} \gamma_{k}}$. Since subcases $1 \mathrm{~d}_{1}$ and $1 \mathrm{~d}_{2}$ are disjoint, for the term $\gamma_{k}$ under consideration there is only one solution. Namely, if the given subcase ( $1 \mathrm{~d}_{1}$ or $1 \mathrm{~d}_{2}$ ) holds for $\gamma_{k}$, then instead of $\mathfrak{T}$ such form of $\mathfrak{D}_{\text {Sld }}^{*}$ should be accepted as a proof sequence of $b$ which corresponds to that subcase.
4.1.1 Consequently, since it is established in 4.1 that for each possible subcase of Case 1, we are able to construct in an effective way such unique augmentation of $\mathfrak{D}$, viz. $\mathfrak{D}_{\mathrm{C} 1}^{*}$, such that $\mathfrak{D}_{\mathrm{C} 1}^{*}$ is a proof sequence of $b$ and that in $\left.\mathfrak{刃}_{\mathrm{C} 1}^{*} \mathrm{~A}\right|_{\bar{R} 1^{*}, \mathrm{R} 2} \gamma_{k}$, then Case 1 is solved. Moreover, since subcases a-d of Case 1 are obviously mutually disjoint, we can conclude that for the term $\gamma_{k}$ under consideration there is only one solution of Case 1.
4.2 There are three remaining cases, cf., 4.1, which we have to investigate. Namely:

Case 2. $\sigma$ is a term of $\mathbf{A}, \tau$ is a term of $\mathbf{V}_{1}, \tau \approx C \sigma \beta_{h},\left.\{\sigma, \tau\}\right|_{\bar{R} 2} \beta_{h}$ and $\left.\left\{\beta_{h}\right\}\right|_{\overline{R 1} 1} \gamma_{k}$. Case 3. $\sigma$ is a term of $\mathbf{V}_{1}, \tau$ is a term of $\mathbf{A}, \boldsymbol{\tau} \approx C \sigma \beta_{h},\left.\{\sigma, \tau\}\right|_{R^{2}} \beta_{h}$ and $\left\{\beta_{h}\right\}_{\overline{\mathrm{RI}}} \gamma_{k}$.
Case 4. Both $\sigma$ and $\tau$ are terms of $\mathrm{V}_{1}, \tau \approx C \sigma \beta_{h},\left.\{\sigma, \tau\}\right|_{\left.\right|^{2} 2} \beta_{h}$ and $\left.\left\{\beta_{h}\right\}\right|_{\overline{R 1} 1} \gamma_{k}$.
Remark II: It is obvious, that if one of the cases 1 or 3 holds for $\gamma_{k}$, then in $\mathfrak{D}$ there must be two distinct terms such that each of them is a term of $A$ and $\gamma_{k}$ is a consequence of them by R1 and R2. On the other hand, if one of the cases 2 or 4 holds for $\gamma_{k}$, then it follows from the definitions of these cases that in $\mathfrak{P}$ there can be only one term such that it is a term of $A$ and $\gamma_{k}$ is a consequence of it by Rl and R2. Hence, in our proof it is not excluded as a possibility that $A=\left\{a_{1}\right\}$.

Using reasonings entirely analogous to these which are presented above we can prove without any difficulty that the cases 2,3 , and 4 can be solved always in a similar way as Case 1. Namely, if one of these cases holds for $\gamma_{k}$, then we are able to construct in an effective way the unique augmentation of $\mathfrak{D}$ such that this augmentation is a proof sequence of $b$ in which $\left.A\right|_{\mathbb{R 1 *}, R 2} \gamma_{k}$. Therefore, since the cases 1-4 are mutually disjoint and
only one of them holds for $\gamma_{k}$ under consideration，we can conclude that if in $\mathfrak{T}$ its subsequence $\mathbf{E}$ is not empty and $\gamma_{k}, 1 \leqslant k \leqslant q$ ，is an arbitrary term of $E$ ，then there is the unique augmentation of $\mathfrak{P}$ such that this augmentation which can be constructed in an effective way is a proof sequence of $b$ in which $A t_{\bar{R} 1^{*}, R_{2}} \gamma_{k}$ ．

4．3 Since in sections 4.1 and 4.2 it is assumed that $\gamma_{k}, 1 \leqslant k \leqslant q$ ，is an arbitrary term of $E$ ，it is self－evident that if we shall apply the methods of a proof which was presented in those sections consecutively to each term of $E$ ，then finally we shall obtain in an effective way the unique augmentation of $(2$ such that this augmentation will be a proof sequence of $b$ in which $\left.A\right|_{R 1^{*}, R_{2}} E$ ．More precisely：
4．3．1 Let us assume that $\mathbf{E}$ is not empty．Since，$c f ., 3.3, \mathbf{E}=\left\{\gamma_{1}, \ldots, \gamma_{q}\right\}$ ， $1 \leqslant q<w<\infty, \mathrm{E}$ is a finite sequence．Then，define

Df． 1 For any $n=1,2,3, \ldots, q$ ：

$$
\mathbf{E}_{\gamma_{n}}=\left\{\begin{array}{l}
\mathbf{E}_{\gamma_{1}}=\left\{\gamma_{2}, \ldots, \gamma_{q}\right\}, \text { i.e., } \mathbf{E}_{\gamma_{1}} \text { is } \mathbf{E} \text { from which } \gamma_{1} \text { is removed, } \\
\mathbf{E}_{\gamma_{m}}=\left\{\gamma_{m+1}, \ldots, \gamma_{q}\right\}, \text { i.e., } \mathbf{E}_{\gamma_{m}} \text { is } \mathbf{E}_{\gamma_{m-1}} \text { from which } \gamma_{m} \text { is removed. }
\end{array}\right.
$$

Since $q$ is finite，it follows at once from Df． 1 that $\mathbf{E}_{\gamma_{q}}$ is the empty sequence．

Now，in the same manner as in 4.1 and 4.2 we construct in an effective way the unique augmentation of $\mathfrak{D}$ in regard to the first term of $\mathbf{E}$ ，viz．$\gamma_{1}$ ． Let us indicate this augmentation by： $\boldsymbol{O}_{\gamma_{1}}^{*}=\left\{\mathbf{A} ; \mathrm{V}_{1 \gamma_{1}}^{*} ; \mathrm{V}_{2 \gamma_{1}}^{*} ; \mathbf{E}_{\gamma_{1}} ; \boldsymbol{F}\right\}$ where $\mathrm{V}_{1 \gamma_{1}}^{*}$ and $\mathbf{V}_{2 \gamma_{1}}^{*}$ are respectively $\mathbf{V}_{1}$ and $\mathbf{V}_{2}$ augmented in regard to $\gamma_{1}, \mathbf{E}_{\gamma_{1}}=$ $\left\{\gamma_{2}, \ldots, \gamma_{q}\right\}$ ．Thus， $\mathfrak{D}_{\gamma_{1}}^{*}$ is a proof sequence of $\boldsymbol{b}$ in which $\left.\mathbf{A}\right|_{\overline{R 1 *}, R_{2}} \gamma_{1}$ ．Since $\boldsymbol{Q}_{\gamma_{1}}^{*}$ is a proof sequence of $\mathbf{b}$ ，we can obtain its augmentation in regard to $\gamma_{2}$ ， viz． $\mathfrak{D}_{\gamma_{2}}^{*}=\left\{\mathbf{A} ; \mathbf{V}_{1 \gamma_{2}}^{*} ; \mathbf{V}_{2 \gamma_{2}}^{*} ; \mathbf{C}_{\gamma_{2}} ; \boldsymbol{F}\right\}$ ．Clearly， $\mathfrak{D}_{\gamma_{2}}^{*}$ is a proof sequence of b in which $\left.\mathbf{A}\right|_{\overline{R 1} \mathbb{k}^{*}, \mathrm{R} 2}\left\{\gamma_{1}, \gamma_{2}\right\}$ ．Applying consecutively the preceding method to all the terms of $E$ according to their order we obtain a finite sequence $\mathfrak{C}=\left\{\mathfrak{D} ; \mathfrak{D}_{\gamma_{1}}^{*} ; \ldots ; \mathfrak{D}_{\gamma_{q}}^{*}\right\}$ containing $q+1$ terms and such that its first term is （2）and if $\sigma_{n}, 2 \leqslant n \leqslant q+1$ ，is a term of $\boldsymbol{C}$ ，then $\sigma_{n}$ is an augmentation of the term $\sigma_{n-1}$ such that $\sigma_{n}$ is a proof sequence of b in which $\left.\mathbf{A}\right|_{\mathbb{R}^{*}, R_{2}}\left\{\gamma_{1}, \ldots, \gamma_{n-1}\right\}$ ． Obviously，the last term of $\mathbb{C}$ ，i．e．， $\mathfrak{D}_{\gamma_{q}}^{*}=\left\{\mathbf{A} ; \mathrm{V}_{1 \gamma_{q}}^{*} ; \mathrm{V}_{2 \gamma_{q}}^{*} ; \mathrm{E}_{\gamma_{q}} ; \boldsymbol{F}\right\}$ ，is a proof sequence of $\mathbf{b}$ in which $\mathbf{E}_{\gamma_{q}}$ is empty，and in which each term of $\mathbf{E}$ is a term of $V_{2 \gamma_{q}}^{*}$ and $\left.A\right|_{R 1^{*}, R_{2}} \mathbf{E}$ ．

Thus，it has been proved in this section that if $\mathfrak{P}=\left\{\mathbf{A} ; \mathbf{V}_{1} ; \mathbf{V}_{2} ; \mathbf{E} ; \mathbf{F}\right\}$ is a proof sequence of $b$ ，and in $\mathfrak{D}$ its subsequence $E$ is not empty，then there is the unique augmentation of D ，viz． $\mathrm{D}_{\gamma_{q}}^{*}$ ，such that $\mathrm{D}_{\gamma_{q}}^{*}$ is a proof sequence of $b$ in which $\left.A\right|_{R 1^{*}, R_{2}} E$ ．

4．4 Since in $\mathfrak{刃}_{\gamma_{q}}^{*} \mathbf{E}_{\gamma_{q}}$ is empty， $\mathfrak{D}_{\gamma_{q}}^{*}=\left\{\mathbf{A} ; \mathbf{V}_{1 \gamma_{q}}^{*} ; \mathbf{V}_{2 \gamma_{q}}^{*} ; \mathbf{F}\right\}$ ．And，in order to simplify this rather cumbersome notation，instead of $\mathfrak{刃}_{\gamma_{q}}^{*}, \mathbf{V}_{1_{\gamma_{q}}}^{*}$ ，and $\mathbf{V}_{2 \gamma_{q}}^{*}$ we shall use $\mathfrak{T}_{0}, \mathrm{~V}_{1 \mathrm{E}}$ ，and $\mathrm{V}_{2 \mathrm{E}}$ ，respectively．Thus， $\mathfrak{D}_{0}=\left\{\mathbf{A} ; \mathrm{V}_{1 \mathrm{E}} ; \mathrm{V}_{2 \mathrm{E}} ; \mathbf{F}\right\}$ will mean the same as $\mathfrak{D}_{\gamma_{q}}^{*}=\left\{\mathbf{A} ; \mathbf{V}_{1 \gamma_{q}}^{*} ; \mathbf{V}_{2 \gamma_{q}}^{*} ; \mathbf{F}\right\}$ ．
5 Now，let us assume that in $\mathfrak{P}$ its subsequence $\mathbf{D}, c f$ ．，3．2．1，is not empty．

Whence, $\mathbf{D}=\{\mathbf{E} ; \mathbf{F}\}, c f$., 3.3 and 3.4 , and, therefore, $\mathfrak{B}=\left\{\mathbf{A} ; \mathbf{V}_{1} ; \mathbf{V}_{2} ; \mathbf{E} ; \mathbf{F}\right\}$ in which at least one of the subsequences, $E$ or $F$, must be unempty. Hence, if $\mathbf{F}$ is empty, then $\mathfrak{T}=\left\{\mathbf{A} ; \mathbf{V}_{1} ; \mathbf{V}_{2} ; \mathbf{E}\right\}$. In such a case, $c f ., 4$, we are always able to replace $\mathfrak{D}$ by its augmentation $\mathfrak{D}_{0}=\left\{\mathbf{A} ; \mathbf{V}_{1 \mathbf{E}} ; \mathbf{V}_{2} \mathbf{E}\right\}$ such that $\mathfrak{D}_{0}$ is a proof sequence of $b$ in which $A{\overleftarrow{R 1 *}, R_{2}}^{b}$. Thus, if in $\boldsymbol{B}$ its subsequence $F$ is empty, Theorem A is proved.
5.1 Therefore, let us assume that in $\mathfrak{T}$ its subsequence $F$ is not empty. Then, if in $\mathfrak{P}, \mathbf{E}$ is empty, $\mathfrak{P}=\left\{\mathbf{A} ; \mathbf{V}_{1} ; \mathbf{V}_{2} ; \mathbf{F}\right\}$. On the other hand, if in $\mathfrak{P}, \mathbf{E}$ is not empty, then $\mathfrak{D}=\left\{\mathbf{A} ; \mathbf{V}_{1} ; \mathbf{V}_{\mathbf{2}} ; \mathbf{E} ; \mathbf{F}\right\}$ and, therefore, $c f$. , 4, we are always able to replace $\mathfrak{O}$ by its augmentation $\mathfrak{O}_{0}=\left\{\mathbf{A} ; \mathrm{V}_{1 \mathbf{E}} ; \mathrm{V}_{2 \mathrm{E}} ; \mathbf{F}\right\}$ such that $\mathfrak{D}_{0}$ is a proof sequence of $b$ in which $\left.A\right|_{\overline{R 1 *}, R 2} E$. Since it is self-evident that $\mathfrak{D}$, in which $E$ is empty but $F$ is not empty, is a particular instance of $\mathfrak{O}_{0}$, in the future only $\mathfrak{T}_{0}$ will be investigated.

Remark III: In order to avoid misunderstanding and confusion it should be noted that if $\mathfrak{D}^{*}$ is an arbitrary augmentation of $\mathfrak{T}$ such that $\mathfrak{D}^{*}$ is a proof sequence of $b$, then the subsequences $V_{1}^{*}$ and $V_{2}^{*}$ of $\mathfrak{D}^{*}$ are always defined in exactly the same way as $\mathbf{V}_{1}$ and $\mathbf{V}_{2}, c f ., 3.1$ and 3.2 , but, obviously, their definitions are automatically adjusted to $\mathfrak{\vartheta}^{*}$. Hence, e.g., in $\left.\mathfrak{刃}_{0} A\right|_{R 1^{*}, R_{2}}\left\{\mathbf{V}_{1 E}\right.$; $\left.V_{2} \mathrm{E}\right\}$.

### 5.2 Assume that in $\mathfrak{D}_{0}$

$$
\mathbf{F}=\left\{\mathbf{f}_{1}, \ldots, \mathbf{f}_{t}\right\}, 1 \leqslant t<w, \mathbf{f}_{t} \approx \mathbf{b} .
$$

Since in $\mathfrak{D}_{0}, \mathbf{E}$ is empty, it follows from the definition of $\mathbf{F}, c f ., 3.4$, that in $\mathfrak{T}_{0}$ there are two terms $\kappa$ and $\lambda$ such that both $\kappa$ and $\lambda$ precede $\mathbf{f}_{1}$, i.e., the first term of $F$, and at least one of them must be a term of $V_{2 E}$ and $\left.\{\kappa, \lambda\}\right|_{\overline{R 2}} \mathbf{f}_{1}$. Whence, clearly, if $\mathbf{F}=\left\{\mathbf{f}_{1}\right\}$, then $\mathbf{f}_{1} \approx \boldsymbol{b}$ and, therefore, in $\mathfrak{D}_{0}$, $\left.A\right|_{R^{*}, R_{2}} b$. Since in such a case Theorem $A$ is proved, let us assume that $\mathbf{F} \neq\left\{\mathbf{f}_{1}\right\}$. Consequently, cf., 3.4, if in $\mathfrak{D}_{0} \mathbf{f}_{k}, 2 \leqslant k \leqslant t$, is a term of $\mathbf{F}$, then either
(1) in $\mathbf{F}$ there is a term $\lambda$ such that $\lambda$ precedes $\mathbf{f}_{k}$ and $\left.\{\lambda\}\right|_{\overline{R 1}} \mathbf{f}_{k}$,
or
(2) in $\mathfrak{T}_{0}$ there are two terms $\mu$ and $\nu$ such that both $\mu$ and $\nu$ precede $\mathbf{f}_{k}$, and at least one of them is either a term of $V_{2 E}$ or a term of $F$, and $\left.\{\mu, \nu\}\right|_{\bar{R} 2} f_{k}$.
5.3 Now, we introduce the following two definitions:

Df. 2 For any $n=1,2,3, \ldots<\infty$ :
$F_{n}=\left\{\begin{aligned} & F_{1}= F \text { from which its first term } f_{1} \text { and every other term, if any, } \\ & \text { which is a consequence of } f_{1} \text { by } R 1 \text { are removed. }\end{aligned}\right\} \begin{aligned} & F_{m}= F_{m-1} \text { from which its first term and every other term, if any, } \\ & \text { which is a consequence of this first term by R1 are removed. }\end{aligned}$
Df. 3 For any $n=1,2,3, \ldots<\infty$ :

$$
S_{n}=\left\{\begin{array}{l}
\mathbf{S}_{1}=\mathbf{F}-\mathbf{F}_{1} \\
\mathbf{S}_{m}=\mathrm{F}_{m-1}-\mathrm{F}_{m} .
\end{array}\right.
$$

Thus, for any $k, 1<k<n$, since $\mathbf{F}_{k}$ and $\mathbf{S}_{k}$ are disjoint, $\mathbf{F}_{k-1}=\left\{\mathbf{S}_{k} ; \mathbf{F}_{k}\right\}$ where $F_{k}$ can be empty. And, if $F_{k}$ is unempty, then $F_{k-1}=\left\{\mathbf{S}_{k} ; \mathbf{S}_{k+1} ; \mathrm{F}_{k+1}\right\}$ and so forth. In order to have a convenient notation in the future we use for an arbitrary $k, 1 \leqslant k \leqslant n, \mathbf{S}_{k}=\left\{\mathbf{s}_{1}^{k}, \ldots, \mathbf{s}_{x}^{k}\right\}, 1 \leqslant x \leqslant t$, assuming that if for $j, 1 \leqslant j \leqslant n, \mathbf{S}_{j} \neq \mathbf{S}_{k}$, then $\mathbf{S}_{j}$ and $\mathbf{S}_{k}$ can have different numbers of terms. This convention will not lead to any misunderstanding in our further deductions.

Remark IV: It follows at once from our assumption concerning the structures of the sequences under investigation, cf., point (c), in section 1, that if $\mathbf{s}_{j}^{k}, 2 \leqslant j \leqslant x$, is a term of $\mathbf{S}_{k}, 1 \leqslant k \leqslant n$, then in $\left.\mathbf{S}_{k}\left\{\mathbf{s}_{1}^{k}\right\}\right|_{R 1} \mathbf{s}_{j}^{k}$.
5.4 Let us assume that for the given $k, 1<k<n$, it was already proved that in $\mathfrak{T}_{0}$ its subsequence $\mathbf{F}=\left\{\mathbf{S}_{1} ; \ldots ; \mathbf{S}_{k-1} ; \mathbf{S}_{k} ; \mathbf{F}_{k}\right\}$ and, moreover, suppose that in $F$ its subsequence $F_{k}$ is not empty. Then, the first term of $F$, say $\mathbf{S}_{1}^{k+1}$, cannot be a consequence by R1 of any term of $\mathfrak{D}_{0}$ which precedes it, since otherwise $\boldsymbol{s}_{1}^{k+1}$ would be a term of one of the subsequences $\mathrm{V}_{1 \mathbf{E}}, \mathrm{~V}_{2} \mathbf{E}$, $\mathbf{S}_{1}, \ldots, \mathbf{S}_{k-1}, \mathbf{S}_{k}$, contrary to the definition of $\mathbf{F}_{k}, c f$. . Remark IV, 3.1, 3.2, 3.3, 4.4, Remark III, 5.1 and 5.3. Hence in $\mathfrak{T}_{0}$ there must be two terms $\mu$ and $\nu$ such that both $\mu$ and $\nu$ precede $\mathbf{s}_{1}^{k+1}$ and $\left.\{\mu, \nu\}\right|_{\overline{R 2}} \mathbf{s}_{1}^{k+1}$. Therefore, in accordance with definitions Df. 2 and Df. 3 in $\mathbf{F}_{k}, \mathbf{s}_{1}^{k+1}$ generates a new subsequence, viz., $\mathbf{S}_{k+1}$. And, since by assumption $\mathbf{F}_{k}$ is unempty, $\mathbf{F}_{k}=$ $\left\{\mathbf{S}_{k+1}, \mathbf{F}_{k+1}\right\}$. Therefore, we proved that: For an arbitrary $k, 1<k<n$, if $\mathfrak{D}_{0}=\left\{\mathbf{A} ; \mathbf{V}_{1 \mathbf{E}} ; \mathbf{V}_{2 \mathbf{E}} ; \mathbf{S}_{1} ; \ldots ; \mathbf{S}_{k} ; \mathbf{F}_{k}\right\}$, then $\mathfrak{D}_{0}=\left\{\mathbf{A} ; \mathbf{V}_{1 \mathbf{E}} ; \mathbf{V}_{2 \mathbf{E}} ; \mathbf{S}_{1} ; \ldots ; \mathbf{S}_{k} ; \mathbf{S}_{k+1} ;\right.$ $\left.\mathrm{F}_{k+1}\right\}$. This statement, together with the facts that $\mathfrak{D}_{0}$ is finite and that in $\mathfrak{D}_{0} \mathbf{S}_{1}$ is not empty, cf., 5.2 and Df. 3, allows us to conclude by an elementary induction that for a certain finite $y, 1 \leqslant y \leqslant t$,

$$
\mathfrak{D}_{0}=\left\{\mathbf{A} ; \mathrm{V}_{1 \mathbf{E}} ; \mathrm{V}_{2 \mathbf{E}} ; \mathbf{S}_{1}, \ldots, \mathbf{S}_{y}\right\}, \mathbf{s}_{z}^{y} \approx \mathbf{b}
$$

where for an arbitrary $\mathbf{S}_{k}, 1 \leqslant k \leqslant y, \mathbf{S}_{k}$ is not empty.
5.5 It follows from the definitions of $\mathbf{F}$ and $\mathrm{S}_{n}$ and the fact, $c f ., 5.4$, that in $\mathfrak{D}_{0} \mathbf{F}=\left\{\mathbf{S}_{1} ; \ldots ; \mathbf{S}_{y}\right\}, 1 \leqslant y \leqslant t$, that for $\mathbf{S}_{k}, 1 \leqslant k \leqslant y$, in $\mathfrak{D}_{0}$ there must be two terms $\sigma$ and $\tau$ such that both $\sigma$ and $\tau$ precede $\mathbf{s}_{1}^{k}$, i.e., the first term of $\mathbf{S}_{k}$ and $\left.\{\sigma, \tau\}\right|_{\mathbb{R}_{2}} \mathbf{s}_{1}^{k}$. Since $\mathfrak{T}_{0}=\left\{\mathbf{A} ; \mathbf{V}_{1 \mathbf{E}} ; \mathbf{V}_{2 \mathbf{E}} ; \mathbf{S}_{1} ; \ldots ; \mathbf{S}_{y}\right\}, \mathfrak{D}_{0}$ is a sequence of $y+3$ mutually disjoint subsequences. Hence, since $\sigma$ and $\tau$ can be the terms of the arbitrary subsequences of $\boldsymbol{\vartheta}_{0}$ which precede $\boldsymbol{S}_{k}$ and they can even belong to two different subsequences, in $\mathfrak{T}_{0}$ there are many possible combinations such that each of them can be eventually the actual instance which satisfies $\left.\{\sigma, \tau\}\right|_{R_{2}} \mathbf{s}_{1}^{k}$. In the future we shall call such possibilities in regard to $S_{k}$ the generic cases of $S_{k}$. Since for our further deductions it is important to know the exact number of the generic cases for each $\mathbf{S}_{h}$, $1 \leqslant h \leqslant y$, this problem will be investigated below.
5.5.1 Clearly, for an arbitrary $S_{h}, 1 \leqslant h \leqslant y$, if in $\mathfrak{D}_{0}$ there are two terms $\sigma$ and $\tau$ such that both $\sigma$ and $\tau$ precede $\mathbf{s}_{1}^{h}$, i.e., the first term of $\mathbf{S}_{h}$ and $\left.\{\sigma, \tau\}\right|_{\overline{R 2}} \mathbf{s}_{1}^{h}$, then the following generic cases ( $\alpha$ ) both $\sigma$ and $\tau$ are the terms of $\mathbf{A} ;(\beta) \sigma$ is a term of $\mathbf{A}$ and $\tau$ is a term of $\mathbf{V}_{1 \mathbf{E}} ;(\gamma) \sigma$ is a term of $\mathbf{V}_{\mathbf{1}} \mathbf{E}$ and $\tau$ is a term of A ; and ( $\delta$ ) both $\sigma$ and $\tau$ are the terms of $\mathrm{V}_{1 \mathrm{E}}$; are impossible,
since otherwise $\mathbf{s}_{1}^{h}$ would be a term of $\mathbf{V}_{2 \mathbf{E}}, c f$., 3.2. Therefore, at least one of the terms, $\sigma$ or $\tau$, must be a term of $\mathbf{V}_{2 \mathrm{E}}$ or a term of $\mathbf{S}_{f}, 1 \leqslant f<h$. Thus, if $h=1$, i.e., $\mathbf{S}_{h}=\mathbf{S}_{1}$, and $\left.\{\sigma, \tau\}\right|_{R^{2}} \mathbf{s}_{1}^{1}$, there are five and only five generic cases of $\mathbf{S}_{1}$, namely:
(a) $\sigma$ is a term of A and $\tau$ is a term of $\mathrm{V}_{2 \mathrm{E}}$,
(b) $\sigma$ is a term of $\mathrm{V}_{1 \mathrm{E}}$ and $\tau$ is a term of $\mathrm{V}_{2 \mathrm{E}}$,
(c) $\sigma$ is a term of $\mathbf{V}_{2 \mathbf{E}}$ and $\tau$ is a term of $\mathbf{A}$,
(d) $\sigma$ is a term of $V_{2 E}$ and $\tau$ is a term of $V_{1 E}$,
(e) Both $\sigma$ and $\tau$ are the terms of $\mathrm{V}_{2 \mathrm{E}}$.
5.5.2 It is self-evident that these five generic cases of $S_{1}$ are also the generic cases of any $S_{h}, 1<h \leqslant y$. But, since in $\boldsymbol{D}_{0}$ the number of subsequences which precede such $S_{h}$ is bigger than the number of subsequences which precede $S_{1}$, there are additional generic cases of $S_{h}$. Thus, for example, if $h=2$, i.e., $\boldsymbol{S}_{h}=\mathbf{S}_{2}$, then since in $\mathfrak{T}_{0}, \mathbf{S}_{2}$ is preceded by $\mathbf{A}$, $\mathrm{V}_{1} \mathrm{E}, \mathrm{V}_{2 \mathrm{E}}$, and $\mathbf{S}_{1}$, there are seven new generic cases of $\mathbf{S}_{2}$, viz. ( $\alpha$ ) $\sigma$ is a term of $\mathbf{A}$, or of $\mathbf{V}_{1} \mathbf{E}$, or of $\mathbf{V}_{2 \mathbf{E}}$ and $\tau$ is a term of $\mathbf{S}_{1} ;(\beta) \sigma$ is a term of $\mathbf{S}_{1}$ and $\tau$ is a term of $\mathbf{A}$, or of $\mathrm{V}_{1 \mathbf{E}}$, or of $\mathrm{V}_{2 \mathrm{E}} ;(\gamma)$ both $\sigma$ and $\tau$ are the terms of $\mathbf{S}_{1}$; such that in $\left.\mathbf{S}_{2}\{\sigma, \tau\}\right|_{R^{2}} \mathbf{S}_{1}$. Thus, there are 12 generic cases of $\mathbf{S}_{2}$. Similarly, there are 21 generic cases of $S_{3}, 32$ of $S_{4}, 45$ of $S_{5}$ and so forth.
5.5.3 The discussion presented above enables us to establish the following formula:

Formula © For any $h, 1 \leqslant h \leqslant y$, if $\mathbf{S}_{h}$ is a subsequence of $\boldsymbol{\mathfrak { D }}_{0}$, then there are $h^{2}+4 h$ generic cases of $\mathrm{S}_{h}$.

We prove Formula © as follows:
(1) If for the given $m, 1 \leqslant m<y, \mathbf{S}_{m+1}$ is a subsequence of $\mathfrak{T}_{0}$, then in $\boldsymbol{\mathcal { P }}_{0}$ there are $3+m$ subsequences which precede $\mathbf{S}_{m+1}$. Hence, it is self-evident that the number of all new generic cases of $S_{m+1}$ is:

$$
((3+m)-1)+((3+m)-1)+1=2 m+5
$$

Using the formula obtained above we define the following function:
Df. 4 For any $n=0,1,2,3, \ldots$

$$
\varphi_{n}=\left\{\begin{array}{l}
\varphi_{0}=5 \\
\varphi_{m}=\varphi_{m-1}+2
\end{array}\right.
$$

Clearly, the value of $\varphi_{0}$ is the number of generic cases of $S_{1}$ and the values of $\varphi_{1}, \varphi_{2}, \varphi_{3}, \ldots$ are respectively the numbers of the new generic cases of $\mathbf{S}_{2}, \mathbf{S}_{3}, \mathbf{S}_{\mathbf{4}}, \ldots$.
(2) It follows from 5.5.1, 5.5.2, and point (1) that for any $f$ and $g, 1 \leqslant f<$ $g<n$, all generic cases of $\mathbf{S}_{f}$ are also the generic cases of $\mathbf{S}_{g}$, but besides them there are $2(g-1)+5$ new generic cases of $\mathbf{S}_{g}$. Then, the function defined in (1) allows us to calculate, for the given finite $m$, the number of all generic cases of $\mathbf{S}_{m}$ as a finite series of $\varphi_{n}$ containing $m$ components. Namely:

$$
\begin{aligned}
\sum_{n=0}^{n=m-1} \varphi_{n} & =\varphi_{0}+\varphi_{1}+\varphi_{2}+\varphi_{3}+\ldots+\varphi_{m-2}+\varphi_{m-1} \\
& =5+7+9+11+\ldots+(2(m-2)+5)+(2(m-1)+5) \\
& =5 m+2+4+6+\ldots+2(m-2)+2(m-1) \\
& =5 m+2(1+2+3+\ldots+(m-2)+(m-1)) \\
& =5 m+2\left(\frac{(m-1) m}{2}\right)=5 m+\left(m^{2}-m\right)=m^{2}+4 m .
\end{aligned}
$$

Thus, for the given finite $m$ Formula © is established.
(3) It remains to prove by induction that for every $\mathrm{S}_{h}, 1 \leqslant h \leqslant y$, Formula © holds. Since we have
(a) Formula © holds for $\mathbf{S}_{1}$.
(b) Assume that for the given $k, 1<k<y$, Formula © holds for $\boldsymbol{S}_{k}$. Hence, $c f$., point (1), the number of all generic cases of $S_{k+1}$ is: $\left(k^{2}+4 k\right)+(2 k+5)=$ $\left(k^{2}+2 k+1\right)+(4 k+4)=(k+1)^{2}+4(k+1)$. Therefore, Formula © holds for $\mathrm{S}_{k+1}$.

The proof of Formula © is complete.

## REFERENCE

[1] Church, A., Introduction to Mathematical Logic, volume 1, Princeton University Press, Princeton, 1956.

