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## ON THE DEPENDENCE OF THE SOLUTION OF GENERALTZED BOLTZMANN EQUATIONS ON THE SCATTERING CROSS SECTION: THE INVERSE PROBLEM Bernd Wiesen

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# On the Dependence of the Solution of Generalized Boltzmann Equations on the Scattering Cross Section: The Inverse Problem 

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#### Abstract

In this paper we consider the problem of the reconstruction of the scattering cross section from the solution of a generalized Boltzmann equation.


Key words: Generalized Boltzmann Equation, Differential Equations in Banach Spaces, Inverse Problems

## 11 Introduction: The Boltzmann equation and scattering cross mections

The evolution of the distribution function of a spatially homogeneous gas consisting of molecules with internal energy is given by:
$\frac{\partial}{\partial t} f\left(t, v, \varepsilon_{1}\right)=J(\sigma, f, f)\left(t, v, \varepsilon_{1}\right) \quad$ with initial value $f_{o} \in L_{1}\left(\mathbb{R}^{3} x \mathbb{R}_{+}\right)$.
In (1.1) we have used the following notations:
$J(\sigma, f, g)=\frac{1}{2} \int_{\Pi^{\prime}} \sqrt{1-\mathrm{e}_{1^{\prime}}-\mathrm{e}_{2}^{\prime}} \sigma\left(E, \mathrm{e}_{1}, \mathrm{e}_{2}, \mathrm{e}_{1}^{\prime}, \mathrm{e}_{2}^{\prime}, \eta \cdot \eta^{\prime}\right)\left[\mathrm{f}^{\prime} \mathrm{g}^{\prime}+\mathrm{f}_{*}^{\prime} \mathrm{g}^{\prime}-\mathrm{fg}_{*}-\mathrm{f}_{* g}\right] \mathrm{d} \mu$
with :
$\Pi=\mathbb{R}^{3} \times \mathbb{R}_{+} x \Delta_{1} \times S_{2}$ with $\Delta_{1}=\left\{\left(e_{1}^{\prime}, e_{2}{ }^{\prime}\right): 0 \leq e_{1}^{\prime}, 0 \leq e_{2}^{\prime}\right.$ and $\left.e_{1}{ }^{\prime}+e_{2}^{\prime} \leq 1\right)$,
$\mathrm{d} \mu=\mathrm{d} \Omega\left(\eta^{\prime}\right) \mathrm{de}_{1}{ }^{\prime} \mathrm{de}_{2}^{\prime} \mathrm{d} \varepsilon_{2} \mathrm{~d} w$,
$E=\frac{1}{2}|v-w|^{2}+\varepsilon_{1}+\varepsilon_{2}, C^{\prime}=\sqrt{2 E\left(1-e_{1}^{\prime}-e_{2}\right)}$ and $e_{i}=\varepsilon_{1} / E, i=1,2$.
$v^{\prime}=\frac{1}{2}\left(v+w+\eta^{\prime} c^{\prime}\right), \quad \varepsilon_{1}^{\prime}=e_{1}^{\prime} E$,
$w^{\prime}=\frac{1}{2}\left(v+w-\eta^{\prime} c^{\prime}\right), \quad \varepsilon_{2}^{\prime}=e_{2}{ }^{\prime} E$,
$f^{\prime}=f\left(t, v^{\prime}, \varepsilon_{1}^{\prime}\right), \quad f^{\prime}{ }_{*}=f\left(t, w^{\prime}, \varepsilon_{2}{ }^{\prime}\right), \quad f_{*}=f\left(t, w, \varepsilon_{2}\right)$

The function $\sigma$ in (1.2) is considered to be an element of the function space introduced below. As usual we denote the space of continous functions from a metric space $X$ into a metric space $Y$ by $C(X \rightarrow Y)$

Definition 1.1: The set $S$ of scattering cross sections is the set of all measurable real valued functions $k$ defined on $\mathbb{R}_{+} \times \Delta_{1} \times \Delta_{1} \times S_{2}$ which have the properties:
(i) $k \in C\left(\mathbb{R}_{+} \times \Delta_{1} \rightarrow L_{1}\left(\Delta_{1} \times[-1,1]\right)\right)$
(ii) $k\left(E, e, e^{\prime}, x\right)=k\left(E, e^{\prime}, e, x\right)$ and $k\left(E, e,\left(e_{1}^{\prime}, e_{2}{ }^{\prime}\right), x\right)=k\left(E, e,\left(e_{2}^{\prime}, e_{1}{ }^{\prime}\right),-x\right)$ a.e.
( iii) $e_{1}+e_{2}=1 \Rightarrow k\left(E, e, e^{\prime}, x\right)=0$ a.e.
(iv ) $\|k\|=\sup _{(E, e)} \int_{\Delta_{1} x S_{2}}\left|k\left(E, e, e^{\prime}, x\right)\right| \sqrt{1-e_{i}-e_{2}^{\prime}} d e^{\prime} d x<\infty$
The set of all nonnegative functions in $S$ will be denoted by $S_{+}$. We denote the closed unit ball of $S$ by $B_{1}$, its boundary by $\partial B_{1}$ and the open unit ball by $B_{1}$.

Notation: For any $\sigma \in S$ we denote
$\sigma_{\tau}(E, e)=\int_{\Delta_{1} \times S_{2}} 2 \pi \sigma\left(E, e, e^{\prime}, x\right) \sqrt{1-e_{i}-e_{2}} d e^{\prime} d x$
and we introduce $L_{i, 1}=\left\{f \in L_{1}: \int_{\mathbb{R}^{3} \mathbf{x} \mathbb{R}_{+}}\left(1+|v|^{2}+\varepsilon_{1}\right)\left|f\left(v, \varepsilon_{1}\right)\right| d \varepsilon_{1} d v<\infty\right\}$.

The aim of this paper is to study the possibility to identify the scattering cross section $\sigma$ from the behavior of the solutions of (1.1) in $\mathrm{C}\left(\left[0, t_{o}\right] \rightarrow \mathrm{L}_{1}\right)$. As a first step in this direction we note a scaling property of such solutions. Suppose we have found a solution $f(\cdot, 0)$ of (1.1) in $C\left(\left[0, t_{0}\right] \rightarrow L_{1}\right)$ for some $t_{0}>0$. We define for $\lambda, \mu>0$
$g(t)=\lambda f(\lambda \mu t, \sigma)$
and we get

$$
\begin{align*}
g(t)=\lambda f(\lambda \mu t, \delta) & =\lambda f_{O}+\lambda \int_{o}^{\lambda \mu t} J(\sigma, f(s), f(s)) d s  \tag{1.8}\\
& =\lambda f_{O}+\int_{0}^{t} J(\mu \sigma, \lambda f(\lambda \mu s), f(\lambda \mu s)) d s=g_{o}+\int_{0}^{t} J(\mu \sigma, g(s), g(s)) d s
\end{align*}
$$

which shows, that $g(\cdot)$ solves (1.1) with data $\lambda f_{O}$ and $\mu \sigma$ in $C\left(\left[0, t_{0} / \mu \lambda\right] \rightarrow L_{1}\right)$. Because of this property we assume in the following:

$$
\left\|f_{\mathrm{o}}\right\|=1 \quad \text { and } \quad \sigma \in \partial \mathrm{B}_{1} .
$$

## 2) Some known results

In this section we collect some results which are needed in the sequel. For the proofs of the following two propositions see Wiesen (1991,1 and 1991,2). We first note that we can split the collision operator J in (1.2) into a gain and a loss part:
$J(\sigma, f, g)=G(\sigma, f, g)-V(\sigma, f, g)$

Proposition 2.1: Let o be in S . Then both $\mathrm{G}(\sigma, \cdot, \cdot)$ and $\mathrm{V}(\sigma, \cdot$,$) are mappings$ from $L_{1} \times L_{1}$ into $L_{1}$ and there hold the estimates:
$\|V(\sigma, f, g)\| \leq 2 \pi\|\sigma\|\|f\|\|g\|$ and $\|G(\sigma, f, g)\| \leq 2 \pi\|\sigma\|\|f\|\|g\|$.
Moreover we have for any $\varphi \in C^{b}\left(\mathbb{R}^{3} \mathbf{x} \mathbb{R}_{+}\right), f, g \in \mathcal{L}_{1}\left(\mathbb{R}^{3} \mathbf{x} \mathbb{R}_{+}\right)$:
$\int_{\mathbb{R}^{3} \mathbf{x R}_{+}} \varphi\left(\mathrm{v}, \varepsilon_{1}\right) \mathrm{J}(\sigma, \mathrm{f}, \mathrm{g})\left(\mathrm{v}, \varepsilon_{1}\right) \mathrm{d} \varepsilon_{1} \mathrm{dv}=$

$$
\begin{equation*}
\frac{1}{2} \int_{\mathbb{R}^{3} \times R_{+} \times \Pi} \sqrt{1-\mathrm{e}_{1}^{\prime}-\mathrm{e}_{2}^{\prime}} \sigma\left(\mathrm{E}, \mathrm{e}_{1}, \mathrm{e}_{2}, \mathrm{e}_{1}^{\prime}, \mathrm{e}_{2}^{\prime}, \eta \cdot \eta^{\prime}\right)\left[\varphi_{*}^{\prime}-\varphi\right]\left[\mathrm{fg}_{*}+\mathrm{f}_{*} \mathrm{~g}\right] \mathrm{d} \mu \otimes \mathrm{~d} \varepsilon_{1} \mathrm{dv} \tag{2.2}
\end{equation*}
$$

For any $\sigma$ in $\partial B_{1}$ and $t_{o}<[4 \pi]^{-1}$ there is a unique function $f(\cdot, \sigma) \in$ $C\left(\left[0, t_{0}\right] \rightarrow L_{1}\right)$ which solves (1.1) with data $f_{0}$ and $\sigma$. If we introduce the following sequence $\left\{\mathrm{G}_{\mathbf{n}}(\sigma)\right\}$ of functions
$G_{O}(\sigma)=f_{0}$
$G_{n}(\sigma)=\frac{1}{n} \sum_{\mu=0}^{n-1} J\left(\sigma, G_{n-1-\mu}(\sigma), G_{\mu}(\delta)\right), n \geq 1$,
then we have the estimate: $\left\|G_{n}(\sigma)\right\| \leq[4 \pi\|\sigma\|]^{n}$ and the solution of (1.1) in $\left[-t_{0}, t_{0}\right]$ may be represented as:
$f(t, \sigma)=\sum_{n=0}^{\infty} t^{n} G_{n}(\sigma), t \in\left[-t_{0}, t_{0}\right]$.

If $\sigma$ and $f_{O}$ are nonnegative functions then for any $t_{0}>0$ there is a unique solution $f(\cdot, \sigma) \in \mathbf{C}\left(\left[0, t_{0}\right] \rightarrow L_{1}\right)$ of (1.1) having the property :
$\forall t \geq 0:\|f(t, \sigma)\|=\left\|f_{0}\right\|$.

Definition 2.1: Let $\sigma$ and $f_{0}$ be a nonnegative functions in $S$ and $L_{1}\left(\mathbb{R}^{3} \times R_{+}\right)$ respectively. We define for $h \geq 2 \pi\|\sigma\|$ the operator
$Q_{h}(\sigma, f, g)=J(\sigma, f, g)+\frac{h}{2}\left\{f \int_{\mathbb{R}^{3} x \mathbb{R}_{+}} g\left(w, \varepsilon_{2}\right) d \varepsilon_{2} d w+g \int_{\mathbb{R}^{3} x \mathbb{R}_{+}} f\left(w, \varepsilon_{2}\right) d \varepsilon_{2} d w\right\}$,
and the following sequence $\left\{\mathrm{H}_{\mathrm{n}}(\sigma)\right\}$ of functions
$\mathrm{H}_{\mathrm{O}}(\sigma)=\mathrm{f}_{\mathrm{O}}$
$H_{n}(\sigma)=\frac{1}{n h} \sum_{\mu=0}^{n-1} \mathbf{Q}_{\mathbf{h}}\left(\sigma, H_{n-1-\mu}(\sigma), H_{\mu}(\sigma)\right) \quad$, if $n \geq 1$.

Proposition 2.2: Let $\sigma \in S$ and $f_{O}$ be nonnegative functions. Then each of the functions $H_{n}(\sigma)$ in (2.6) is nonnegative and we have $\left\|H_{n}(\sigma)\right\|=\left\|f_{0}\right\|=1$ for any $n \in \mathbb{N}$. The unique solution of (1.1) can be represented as
$f(t, \sigma)=\sum_{n=0}^{\infty} e^{-h t}\left(1-e^{-h t}\right)^{n} H_{n}(\sigma)$,
where we can choose any $h \geq 2 \pi\|\sigma\|$
Remark: For the rest of this paper we assume $f_{O}$ to be a nonnegative function.

Remark: For both sequences $\left\{G_{n}(\cdot), n \geq 0\right\}$ and $\left\{H_{n}(\cdot), n \geq 0\right\}, f_{o}$ is called the start value of the sequence.

## 3) Injectivity propertles of the solution of (1.1)

Proposition 3.1: Let $f\left(\cdot, \sigma_{1}\right)$ and $f\left(\cdot, \sigma_{2}\right)$ be two solutions of (1.1) with data $\sigma_{1} \in B_{1}^{\prime}$ and $\sigma_{2} \in B_{1}^{\prime}$ and suppose: $f\left(\cdot, \sigma_{1}\right) \neq f\left(\cdot, \sigma_{2}\right)$ in $\mathbf{C}\left(\left[0, t_{0}\right] \rightarrow L_{1}\right)$, where $t_{0}$ is an arbitrary positive time. Then the set $M$ of all times $t \in \mathbb{R}_{+}$with the property: $f\left(t, \sigma_{1}\right)=f\left(t, \sigma_{2}\right)$ in $L_{1}$ is closed and there is no finite limit point in M .

Proof: The closedness of M is a direct consequence of the continuity of $\mathrm{f}\left(\cdot, \sigma_{1}\right)$ and $\mathrm{f}\left(\cdot, \sigma_{2}\right)$, see proposition 2.1.
Assume now, we have a finite limit point $T_{1} \in M$. Then there exists a sequence $\left\{t_{n}\right.$ \} in $M$ with the property:
$\forall \mathrm{n} \in \mathbb{N}: \mathrm{t}_{\mathrm{n}} \neq \mathrm{T}_{1} \quad$ and $\quad \lim _{\mathrm{n} \rightarrow \infty} \mathrm{t}_{\mathrm{n}}=\mathrm{T}_{1}$
We set $\Psi_{0}=f\left(T_{1}, \sigma_{1}\right)=f\left(T_{1}, \sigma_{2}\right)$ and define the sequences $\left\{G_{n}\left(\sigma_{1}\right), n \geq 0\right\}$ and $\left\{\mathrm{G}_{\mathrm{n}}\left(\sigma_{2}\right), \mathrm{n} \geq 0\right\}$ with start value $\Psi_{\mathrm{o}}$. In the time intervall $\left[\mathrm{T}_{1}-[4 \pi]^{-1}, \mathrm{~T}_{1}+[4 \pi]^{-1}\right]$ the solutions of (1.1) with data $\sigma_{1}$ and $\sigma_{2}$ can be represented as
$f\left(t, \sigma_{1}\right)=\sum_{n=0}^{\infty}\left(t-T_{1}\right)^{n} G_{n}\left(\sigma_{1}\right)$ und $f\left(t, \sigma_{2}\right)=\sum_{n=0}^{\infty}\left(t-T_{1}\right)^{n} G_{n}\left(\sigma_{2}\right)$
Now the fact: $\forall \mathrm{n} \in \mathbb{N}: \mathbf{f}\left(\mathrm{t}_{\mathrm{n}}, \sigma_{1}\right)=\mathbf{f}\left(\mathrm{t}_{\mathrm{n}}, \sigma_{2}\right)$ yields:
$\forall k \in \mathbb{N}: \lim _{n \rightarrow \infty} \frac{\left\|f\left(t_{n}, \sigma_{1}\right)-f\left(t_{n}, \sigma_{2}\right)\right\|}{\left|T_{1}-t_{n}\right|^{k}}=0$
and we get with the help of standard techniques from (3.1) and (3.2):
$\forall \mathrm{n} \in \mathbb{N}: \mathrm{G}_{\mathrm{n}}\left(\sigma_{\mathbf{1}}\right)=\mathrm{G}_{\mathrm{n}}\left(\sigma_{\mathbf{2}}\right)$.
But (3.3) implies $f\left(\cdot, \sigma_{1}\right)=\int\left(\cdot, \sigma_{2}\right)$ in $C\left(\left[T_{1}-[4 \pi]^{-1}, T_{1}+[4 \pi]^{-1}\right] \rightarrow L_{1}\right)$. Now an iteration procedure yields $\mathrm{f}\left(\cdot, \sigma_{1}\right)=\mathrm{f}\left(\cdot, \sigma_{2}\right)$ in $\mathrm{C}\left(\left[0, \mathrm{t}_{\mathbf{O}}\right] \rightarrow \mathrm{L}_{1}\right)$ which gives a contradiction.

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$$

An immediate consequence of proposition 3.1 and of the theorem of the continuation of the solution of differential equations in Banach spaces (see e.g. Martin, 1976, chapter 6) is the following

Corollary 3.1: Let $\sigma_{1}$ and $\sigma_{2}$ be two nonnegative functions in $S$ and let $f\left(\cdot, \sigma_{1}\right)$ and $f\left(\cdot, \sigma_{2}\right)$ be the corresponding solutions of (1.1). Then there are equivalent:
(i) There exists a $t_{0}>0$ such that $f\left(\cdot, \sigma_{1}\right)=f\left(\cdot, \sigma_{2}\right)$ in $\mathrm{C}\left(\left[0, t_{0}\right] \rightarrow L_{i}\right)$
(ii) $\forall \mathrm{n} \in \mathbb{N}: \mathrm{G}_{\mathrm{n}}\left(\sigma_{1}\right)=\mathrm{G}_{\mathrm{n}}\left(\sigma_{2}\right)$, where the functions $\mathrm{G}_{\mathrm{n}}(\cdot)$ are given by (2.3) and for any $t_{0}>0$ there holds : $f\left(\cdot, \sigma_{1}\right)=f\left(\cdot, \sigma_{2}\right)$ in $C\left(\left[0, t_{0}\right] \rightarrow L_{1}\right)$.

Theorem 3.1: Let $\sigma_{1}, \sigma_{2} \in S_{+}$be two different scattering cross sections. Let $f\left(\cdot, \sigma_{1}\right)$ and $f\left(\cdot, \sigma_{2}\right)$ be the corresponding solutions of (1.1). We denote $\sigma(\lambda)=\lambda \sigma_{1}+(1-\lambda) \sigma_{2}, 0 \leqslant \lambda \leqslant 1$. Suppose there exists a $t_{1}>0$ such that $f\left(t_{1}, \sigma_{1}\right) \neq f\left(t_{1}, \sigma_{2}\right)$. Then, for any $t_{0}>0$, the map
$[0,1] \ni \lambda \rightarrow f(\cdot, \sigma(\lambda)) \in \mathbf{C}\left(\left[0, t_{0}\right] \rightarrow L_{1}\right)$, with $\sigma(\lambda)=\sigma_{1}+\lambda\left(\sigma_{2}-\sigma_{1}\right)$,
$f(\cdot, \sigma(\lambda))$ being the solution of (1.1) with data $f_{O}$ and $\sigma(\lambda)$, is injective.

Proof: As a consequence of corollary 3.1 there exists a $\mathrm{N}>0$ such that:
$\mathrm{G}_{\mathrm{N}}\left(\sigma_{1}\right) \neq \mathrm{G}_{\mathrm{N}}\left(\sigma_{2}\right)$,
where the functions $G_{n}(\cdot)$ are given by (2.3). We set $M=\inf (n$ : $\left.\mathrm{G}_{\mathrm{n}}\left(\sigma_{1}\right) \neq \mathrm{G}_{\mathbf{n}}\left(\sigma_{2}\right)\right\}$. Using (2.3) we obtain
$\mathrm{G}_{\mathbf{n}}(\sigma(\lambda))=\frac{1}{n} \sum_{\mu=0}^{\mathrm{n}-1}(1-\lambda) J\left(\sigma_{1}, \mathrm{G}_{\mathrm{n}-1-\mu}(\sigma(\lambda)), \mathrm{G}_{\mu}(\sigma(\lambda))\right)$

$$
+\lambda J\left(\sigma_{2}, G_{\mathbf{n}-1-\mu}(\sigma(\lambda)), \mathrm{G}_{\mu}(\sigma(\lambda))\right)
$$

which yields
$G_{n}(\sigma(\lambda))=G_{n}\left(\sigma_{1}\right)=G_{n}\left(\sigma_{2}\right), n=0,1, ., M-1, \quad$ and
$\mathrm{G}_{\mathbf{M}}(\sigma(\lambda))=(1-\lambda) \mathbf{G}_{\mathbf{M}}\left(\sigma_{1}\right)+\lambda \mathrm{G}_{\mathbf{M}}\left(\sigma_{2}\right)$.
Now equation (3.4) implies that we have for any two $\lambda_{1}$ and $\lambda_{2} \in[0,1]$ :
$\lambda_{1} \neq \lambda_{2} \Rightarrow \mathrm{G}_{\mathbf{M}}\left(\sigma\left(\lambda_{1}\right)\right) \neq \mathrm{G}_{\mathbf{M}}\left(\sigma\left(\lambda_{2}\right)\right)$
and the assertion follows from corollary 3.1.

## 4) Separation properties of the collision operator $1(l)$

In this section we use the injectivity criterion of corollary 3.1 to see that there exists for any two different scattering cross sections $\sigma_{1}$ and $\sigma_{2}$ an initial condition $f_{o}$ such that the corresponding solutions of (1.1) are different, regardless of the observation time $t_{0}$. To this end we note that the function $G_{1}$ of the sequence (2.3) is given by $G_{1}(\sigma)=J(\sigma, f, f)$ where $f$ is the start function of the sequence. So all we have to show is that, for any two different $\sigma_{1}$ and $\sigma_{2}$, there is a nonnegative function f such that $J\left(\sigma_{1}, \mathrm{f}, \mathrm{f}\right) \neq \mathrm{J}\left(\sigma_{2}, \mathrm{f}, \mathrm{f}\right)$.

Lemma 4.1: The family of mappings
\{ $\mathbf{S} \boldsymbol{\exists} \sigma \rightarrow \mathrm{J}(\sigma, \mathrm{f}, \mathrm{f})$, f nonnegative function in $\mathrm{L}_{1,1}$ \}
separates on $S$.

Proof: Let $\sigma_{1}$ and $\sigma_{2}$ be two different scattering cross sections. Due to property (i) of definition 1.1 there exist rational numbers $E_{O} \in \mathbb{R}_{+}$and $\mathrm{e}_{0}=\left(\mathrm{e}_{10}, \mathrm{e}_{20}\right)$ \& $\Delta_{1}$ such that
$k\left(E_{0}, e_{0}, .,\right):\left(e^{\prime}, x\right) \rightarrow \sigma_{1}\left(E_{0}, e_{0}, e^{\prime}, x\right)-\sigma_{2}\left(E_{0}, e_{0}, e^{\prime}, x\right)$
is not the null function. We have to show : $\exists f \in \mathrm{~L}_{1,1}: J(k, f, f) \neq 0$ in $\mathrm{L}_{1}$. Suppose this is not the case. Then we have:

$$
\begin{equation*}
\forall \varphi \in C^{b}\left(\mathbb{R}^{3} \mathbf{x} \mathbb{R}_{+}\right), f \in L_{1,1}: \int_{\mathbb{R}^{3} x R_{+}} \varphi\left(v, \varepsilon_{1}\right) J(k, f, f)\left(v, \varepsilon_{1}\right) d \varepsilon_{1} d v=0 \tag{4.2}
\end{equation*}
$$

Due to the properties of the function $k$, recall (i), (ii) and (iii) in definition 1.1, the map

$$
\begin{align*}
\mathbb{R}^{3} \mathbf{x} \mathbb{R}_{+} & \times \mathbb{R}^{3} \mathbf{x} \mathbb{R}_{+} \backslash\left(\left(\mathbf{v}, \varepsilon_{1}, \mathbf{v}, \varepsilon_{2}\right), v \in \mathbb{R}^{3}, \varepsilon_{1}, \varepsilon_{2} \in \mathbb{R}_{+}\right\} \quad \exists\left(\mathrm{v}, \varepsilon_{1}, \mathbf{w}, \varepsilon_{2}\right) \rightarrow \\
& \int_{\Delta_{1} \times S_{2}} \mathrm{k}\left(E, \mathbf{e}, \mathrm{e}^{\prime}, \eta \cdot \eta^{\prime}\right)\left[\varphi\left(\mathbf{v}^{\prime}, \varepsilon_{1}\right)-\varphi\left(\mathrm{v}, \varepsilon_{1}\right)\right] \sqrt{1-\mathrm{e}_{1}^{\prime}-\mathbf{e}_{2}^{\prime}} \mathrm{d} \Omega\left(\eta^{\prime}\right) \mathrm{de} \tag{4.3}
\end{align*}
$$

is a continous function in its variables ( $\mathrm{v}, \varepsilon_{1}, \mathrm{w}, \varepsilon_{2}$ ). Recall that we used in (4.3) the notations (1.3) and (1.4). We set
$\varepsilon_{10}=e_{10} E_{0}, \varepsilon_{20}=e_{20} E_{0}, v_{1}=(r, 0,0), \quad v_{2}=(-r, 0,0), \quad r=\frac{E_{0}\left(1-e_{10}-e_{20}\right)}{2}$
and define for $n \in \mathbb{N}$ :
$\mathrm{g}_{\mathrm{n}}(\mathrm{v}, \varepsilon)=\frac{\mathrm{n}}{2} \sqrt{\frac{\mathrm{n}}{}^{3}}\left[\mathrm{H}\left(\varepsilon-\varepsilon_{10}\right) \exp \left[-\mathrm{n}\left(\left|v-v_{1}\right|^{2}+\varepsilon\right)\right]+H\left(\varepsilon-\varepsilon_{20}\right) \exp \left[-n\left(\left|v-v_{2}\right|^{2}+\varepsilon\right)\right]\right]$,
where $H(\cdot)$ is the Heaviside function. It can be seen easily that for $n \in \mathbb{N}$ $\mathrm{g}_{\mathrm{n}}$ is in $\mathrm{L}_{1,1}$. In addition we have
$\forall n \in \mathbb{N}: \int_{R^{3} x R_{+}} g_{n}\left(v, \varepsilon_{1}\right) d \varepsilon_{1} d v=1$
Because of (4.2) there holds:
$\forall \mathrm{n} \in \mathbb{N}, \varphi \in \mathbb{C}^{\mathrm{b}}\left(\mathbb{R}^{3} \mathbf{x} \mathbb{R}_{+}\right): \int_{\mathbb{R}^{3} \mathbf{x} \mathbb{R}_{+}} \varphi\left(\mathrm{v}, \varepsilon_{1}\right) \mathrm{J}\left(\mathrm{k}, \mathrm{g}_{\mathrm{n}}, \mathrm{g}_{\mathbf{n}}\right)\left(\mathrm{v}, \varepsilon_{1}\right) \mathrm{d} \varepsilon_{1} \mathrm{dv}=0$.
Due to the continuity of the mapping (4.3) we get (see e.g. Folland, 1976, Thm. 0.13):

$$
\begin{align*}
& \lim _{n \rightarrow \infty} \int_{\mathbb{R}^{3} \mathbf{x} \mathbb{R}_{+}} \varphi\left(\mathbf{v}, \varepsilon_{1}\right) \mathrm{J}\left(\mathrm{k}, \mathrm{~g}_{\mathrm{n}}, \mathrm{~g}_{\mathbf{n}}\right)\left(\mathrm{v}, \varepsilon_{1}\right) \mathrm{d} \varepsilon_{1} \mathrm{dv}= \\
& \frac{1}{4} \int_{\Delta_{1} \mathrm{xS}_{2}} \mathrm{k}\left(\mathrm{E}_{\mathrm{O}}, \mathrm{e}_{\mathrm{o}}, \mathrm{e}^{\prime}, \eta_{0} \cdot \eta^{\prime}\right) \sqrt{1-\mathrm{e}_{1}^{\prime}-\mathrm{e}_{2}^{\prime}} \\
& -\left[\varphi\left(\frac{\eta^{\prime}}{2} \sqrt{2 \mathrm{E}_{\mathrm{O}}\left(1-\mathrm{e}_{1}^{\prime}-\mathrm{e}_{2}{ }^{\prime}\right)}, \mathrm{e}_{1} \mathrm{E}_{\mathrm{o}}\right)-\varphi\left(\mathrm{v}_{1}, \varepsilon_{10}\right)\right] \mathrm{d} \Omega\left(\eta^{\prime}\right) \mathrm{d} \mathrm{e}^{\prime} \\
& \frac{1}{4} \int_{\Delta_{1} \times S_{2}} \mathrm{k}\left(\mathrm{E}_{\mathrm{o}}, \mathrm{e}_{\mathrm{o}}, \mathrm{e}^{\prime},-\cdots \eta_{0} \cdot \eta^{\prime}\right) \sqrt{1-\mathrm{e}_{1}^{\prime}-\mathrm{e}_{2}}  \tag{4.5}\\
& -\left[\varphi\left(\frac{\eta^{\prime}}{2} \sqrt{2 E_{0}\left(1-e_{1}{ }^{\prime}-e_{2}{ }^{2}\right)}, e_{1} E_{0}\right)-\varphi\left(v_{2}, \varepsilon_{20}\right)\right] d \Omega\left(\eta^{\prime}\right) d e^{\prime},
\end{align*}
$$

In (4.5) $\eta_{O}$ denots the unit vector in the direction of $v_{1}$. If we use now the symmetry properties of k , see (4.1) and (1.6), we get from (4.4) and (4.5):
$\forall \varphi \in C^{b}\left(\mathbb{R}^{3} \mathbf{x} \mathbb{R}_{+}\right):$

$$
\begin{array}{r}
0=\int_{\Delta_{1} \times S_{2}} \mathrm{k}\left(\mathrm{E}_{\mathrm{O}}, \mathrm{e}_{\mathrm{O}}, \mathrm{e}^{\prime}, \eta_{\mathrm{O}} \cdot \eta^{\prime}\right) \sqrt{1-\mathrm{e}_{1}^{\prime}-\mathrm{e}_{2}^{\prime}} \quad\left[2 \varphi\left(\frac{\eta^{\prime}}{2} \sqrt{\left.2 \mathrm{E}_{\mathrm{O}}\left(1-\mathrm{e}_{1}^{\prime}-\mathrm{e}_{2}\right)^{\prime}\right)} \mathrm{e}_{1} \mathrm{E}_{\mathrm{O}}\right)\right.  \tag{4.6}\\
\left.-\varphi\left(\mathrm{v}_{1}, \varepsilon_{10}\right)-\varphi\left(v_{2}, \varepsilon_{20}\right)\right] \mathrm{d} \Omega\left(\eta^{\prime}\right) \mathrm{de}^{\prime}
\end{array}
$$

To discuss the integral on the right hand side of (4.6) we perform the following changes of integration variables:
$\left(e_{1}^{\prime}, e_{2}^{\prime}\right) \rightarrow\left(z=e_{1}^{\prime}+e_{2}^{\prime}, y=e_{1}^{\prime}-e_{2}^{\prime}\right)$
$y \rightarrow y^{\prime}=\frac{y}{z} \quad, \quad z \rightarrow z^{\prime}=1-z, \quad z \rightarrow r=\sqrt{\frac{E_{0} z^{\prime}}{2}}$,
$y^{\prime} \rightarrow y^{\prime \prime}=-\left(E_{O}-2 r^{2}\right)\left(1+y^{\prime}\right)$
and we obtain:
$0=\int_{G\left(E_{O}\right)} k\left(E_{O},\left(e_{10}, e_{20}\right),\left(e_{1}^{\prime}(x, z), e_{2}^{\prime}(x, z)\right), \eta_{0} \cdot \eta^{\prime}(x)\right) \sqrt{\frac{8}{E_{O}}}$

$$
\begin{equation*}
\cdot\left[\varphi(x, z)-\frac{1}{2}\left(\varphi\left(v_{1}, \varepsilon_{10}\right)-\varphi\left(v_{2}, \varepsilon_{20}\right)\right)\right] d z \otimes d^{3} x \tag{4.7}
\end{equation*}
$$

where we have: $G\left(E_{O}\right)=\left\{(x, z) \in \mathbb{R}^{3} x \mathbb{R}_{+}:|x|^{2} \leq E_{O} / 2\right.$ and $\left.0 \leq z \leq E_{O}-2|x|^{2}\right\}$, $e_{1}^{\prime}(x, z)=\frac{z}{E_{0}}, \quad e_{2}{ }^{\prime}(x, z)=\frac{E_{0}-2|x|^{2}-z}{E_{0}}, \quad \eta^{\prime}(x)=\frac{x}{|x|}$.

We consider the following measures on $\mathrm{G}\left(\mathrm{E}_{\mathrm{O}}\right)$ and its induced Borel $\sigma$ algebra.
$d \mu_{1}(x, z)=\sqrt{\frac{8}{E_{0}^{5}}} k\left(E_{O}, e_{10}, e_{20}, e_{1}^{\prime}(x, z), e_{2}^{\prime}(x, z), \eta_{O}^{\prime} \eta^{\prime}(x)\right) d z \otimes d^{3} x$
and

$$
\begin{align*}
d \mu_{2}(x, z)= & \frac{\pi}{2}\left[\int_{\Delta_{1} x[-1,1]} k\left(E_{O}, e_{o}, e^{\prime}, x\right) \sqrt{1-e_{1}^{\prime}-e_{2}^{\prime}} d x d e^{\prime}\right]  \tag{4.9}\\
& \cdot\left[\delta\left(x-v_{1}\right) \otimes \delta\left(z-\varepsilon_{1 O}\right)+\delta\left(x-v_{2}\right) \otimes \delta\left(z-\varepsilon_{Z O}\right)\right]
\end{align*}
$$

and note that (4.7) implies:

$$
\begin{equation*}
\forall \varphi \in C^{b}\left(\mathbb{R}^{3} x R_{+}\right): \int_{G\left(E_{O}\right)} \varphi(x, z) d \mu_{1}(x, z)-\int_{G\left(E_{O}\right)} \varphi(x, z) d \mu_{2}(x, z)=0 \tag{4.10}
\end{equation*}
$$

Because of (4.1), $d \mu_{1}(x, z)$ is not the zero measure. A comparison of (4.8) and (4.9) shows that $d \mu_{1}$ is absolutely continous with respect to the 4 dimensional Lebesque measure whereas $\mathrm{d} \mu_{2}$ is singular to this measure. So both measures are different.
Now, using that the dual space of the real measures on $G\left(E_{O}\right)$ is isomorphic to $C^{b}\left(G\left(E_{O}\right)\right)$, we get:
$\exists \varphi^{\prime} \in \mathrm{C}^{\mathrm{b}}\left(\mathrm{G}\left(\mathrm{E}_{\mathrm{O}}\right)\right): \int_{\mathbf{G}\left(\mathrm{E}_{\mathrm{O}}\right)} \varphi^{\prime}(\mathrm{x}, \mathrm{z}) \mathrm{d} \mu_{1}(\mathrm{x}, \mathrm{z}) \quad-\int_{\mathbf{G}\left(\mathrm{E}_{\mathrm{O}}\right)} \varphi^{\prime}(\mathrm{x}, \mathrm{z}) \mathrm{d} \mu_{2}(\mathrm{x}, \mathrm{z}) \neq 0$

A comparison of (4.10) and (4.11) shows that we get a contradiction, provided we can show that the restriction map
$R\left(E_{O}\right): C^{b}\left(R^{3} x R_{+}\right) \rightarrow C^{b}\left(G\left(E_{O}\right)\right):\left.\varphi \rightarrow \varphi\right|_{G\left(E_{O}\right)}$
is surjective.
Due to the compactness of $G\left(E_{O}\right)$ there exists a $r>0$ such that
$G\left(E_{O}\right) \subset B_{4}(0, r)=\left\{(x, z) \in \mathbb{R}^{3} x \mathbb{R}_{+}:|x|^{\left.2_{+z} \leq r\right\}}\right.$.
Now Tietze's extension theorem (see Reed-Simon, 1980, Thm. IV.11) yields the surjectivity of the mapping
$R_{r}: C^{b}\left(B_{4}(0, r)\right) \rightarrow C^{b}\left(G\left(E_{O}\right)\right):\left.\varphi \rightarrow \varphi\right|_{G\left(E_{O}\right)}$
The surjectivity of the map $R\left(E_{0}\right)$ is now a direct consequence of Urysohn's lemma (see Rudin, 1970, Thm. 2.12).

Remark: The set $\sqrt{Q_{+}}=\left\{x \in \mathbb{R}: x^{2} \in Q\right\}$ is countable.

Theorem 4.1: There exists a sequence $\left\{f_{n}\right\}$ of nonnegative functions in $L_{1,1}$ with the properties
(i) $\forall n \in \mathbb{N}: \int_{\mathbb{R}^{3} x \mathbb{R}_{+}} f_{n}\left(v, \varepsilon_{1}\right) d \varepsilon_{1} d v \leq 1$
(ii) $\forall n \in N: \int_{\mathbb{R}^{3} x \mathbb{R}_{+}} v f_{n}\left(v, \varepsilon_{1}\right) d \varepsilon_{1} d v=0$
(iii) $\exists \mathbf{C} \in \mathbb{R}_{+}: \forall n \in \mathbb{N}: \int_{\mathbb{R}^{3} \mathbf{x} \mathbb{R}_{+}}\left(1+|v|^{2}+\varepsilon_{1}\right) f_{n}\left(v, \varepsilon_{1}\right) d \varepsilon_{1} d v \leq C$,
such that the set of mappings $\left\{S \ni \emptyset \rightarrow J\left(\varnothing, f_{n}, f_{n}\right) \in L_{i}\right\}$ separates on $S$.

Proof: We start with the sequence $\left\{g_{n}\right\}$ of functions introduced in (4.4). These functions are parametrized by the the values $r, \varepsilon_{10}, \varepsilon_{20}$. Therefore we write here $g_{n}\left(r, \varepsilon_{10}, \varepsilon_{20}\right)(v, \varepsilon)$. We have the following properties:
(a) $\int_{\mathbb{R}^{3} \times \mathbb{R}_{+}} g_{n}\left(r, \varepsilon_{10}, \varepsilon_{20}\right)(v, \varepsilon) d \varepsilon d v=1 \quad$ for arbitrary $n, r, \varepsilon_{10}, \varepsilon_{20}$
(b) $\int_{\mathbb{R}^{3} X_{X}} \vee g_{n}\left(r, \varepsilon_{10}, \varepsilon_{20}\right)(v, \varepsilon) d \varepsilon d v=0 \quad$ for arbitrary $n, r, \varepsilon_{10}, \varepsilon_{20}$
$(\mathrm{c}) \int_{\mathbb{R}^{3} \mathrm{x} \mathrm{R}_{+}}\left(1+|v|^{2}+\varepsilon\right)_{g_{n}}\left(\mathrm{r}, \varepsilon_{10} \varepsilon_{20}\right)(\mathrm{v}, \varepsilon) \mathrm{d} \varepsilon \mathrm{d} v=1+\mathrm{r}^{2}+\frac{1}{2}\left(\varepsilon_{10}+\varepsilon_{2 O}\right)+\frac{5}{2 n}$, because we have $v_{1}=(r, 0,0)$ und $v_{2}=(-r, 0,0)$.

We define for arbitrary but fixed $\mathrm{C}>0$ the functions:

$$
\begin{aligned}
\mathrm{f}_{\mathrm{n}}\left(\mathrm{r}, \varepsilon_{10}, \varepsilon_{20}, \mathrm{C}\right)= & \lambda\left(\mathrm{r}, \varepsilon_{10}, \varepsilon_{20}, C\right) \mathrm{g}_{\mathrm{n}}\left(\mathrm{r}, \varepsilon_{10}, \varepsilon_{20}\right) \quad \text { with } \\
& \lambda\left(\mathrm{r}, \varepsilon_{10}, \varepsilon_{20}, \mathrm{C}\right)=\min \left[1, \frac{C}{r^{2}+\frac{1}{2}\left(\varepsilon_{10}+\varepsilon_{20}\right)+\frac{7}{2}}\right]
\end{aligned}
$$

 countable and the functions $f_{n}$ have the properties (4.12), (4.13) and (4.14). Moreover we can see easily from the proof of lemma 4.1 that the set of mappings
$\left\{S \exists \sigma \rightarrow J(\sigma, f, f) \in L_{1}, \quad f \in M_{S}\right\}$
separates on $S$.

Defintion 4.1: We call a sequence $\left\{h_{n}\right\}$ of nonnegative functions in $L_{1,1}$ a separating sequence, if it has the properties mentioned in theorem 4.1. The constant C occuring in (4.14) is called energy bound of the sequence $\left\{\mathrm{h}_{\mathrm{n}}\right.$ \}.

Notation: As usual we denote the space of the bounded sequences over some Banach space $X$ by $l_{\infty}(X)$. It is equipped with the norm $1_{\infty}{ }^{3} y \rightarrow\|y\|_{\infty}=\sup _{n \in \mathbb{N}}\left\|y_{n}\right\|_{x}$.

Suppose now we have a separating sequence $\left\{h_{n}\right\}$. Then theorem 4.1 motivates the study of the following mapping
$S_{+} \exists \sigma \rightarrow F(\sigma)=\left(f_{1}(\cdot, \sigma), f_{2}(\cdot, \sigma), \cdot,\right) \in 1_{\infty}\left(C\left(\left[0, t_{0}\right] \rightarrow L_{1}\right)\right)$.

In (4.15) $t_{o}$ is any positive time and the functions $f_{i}(\cdot, 0)$ are the solutions of (1.1) with data $\sigma$ and $h_{1}$. It is a direct consequence of theorem 4.1 that this mapping is injective. So we study the inverse of the mapping (4.15) on the image of $S_{+}$which will be denoted by $F\left(S_{+}\right)$.

## 5) On the inverse problem

Lemma 5.1: Let $f_{1}$ and $f_{2}$ be two nonnegative functions in $L_{1}$ with the following property: there are constants $C_{1}, C_{2}>0$ such that:
$\forall R>0: \int_{B_{4} C(R)} f_{i}(v, \varepsilon) d \varepsilon d v<\frac{C_{i}}{1+R^{2}}, i=1,2$,
where we have $B_{4}{ }^{c}(R)=\left\{(v, \varepsilon):|v|^{2}+\varepsilon>R^{2}\right\}$. Let $k \in S_{+}$a function with the property:
$\exists D>0: \forall E \leq 2 D^{2}, e \in \Delta_{1}: k(E, e, \cdot \cdot)=0$.
Then we have :
$\| J\left(k, f_{1}, f_{2} \| \quad \leq \frac{4 \pi\|k\|}{1+D^{2}}\left(C_{1}\left\|f_{2}\right\|+C_{2}\left\|f_{1}\right\|\right)\right.$.

Proof: Using proposition 2.1 and the nonnegativity of $f_{1}, f_{2}$ and $k$ we get

$$
\begin{align*}
\|J(k, f, g)\| \leq & 2 \int_{\mathbb{R}^{3} \times R_{+}} V(k, f, g)\left(v, \varepsilon_{1}\right) d \varepsilon_{1} d v \\
& =\int_{\mathbb{R}^{3} \times R_{+} \times R^{3} \times R_{+} \times \Delta_{1} \times S_{2}} \begin{aligned}
& k\left(E, e e^{\prime}, \eta \cdot \eta^{\prime}\right) \sqrt{1-e_{1}^{\prime}-e_{2}^{\prime}}\left[f_{1}\left(v, \varepsilon_{1}\right) f_{2}\left(w, \varepsilon_{2}\right)\right. \\
& \left.+f_{2}\left(v, \varepsilon_{1}\right) f_{1}\left(w, \varepsilon_{2}\right)\right] d \Omega\left(\eta^{\prime}\right) d e^{\prime} d \varepsilon_{2} d w d \varepsilon_{1} d v .
\end{aligned} \tag{5.4}
\end{align*}
$$

Recall that we have in (5.4):
$E=\frac{1}{2}|v-w|^{2}+\varepsilon_{1}+\varepsilon_{2} \leq|v|^{2}+\varepsilon_{1}+|w|^{2}+\varepsilon_{2}$

The set $\Gamma=\mathbb{R}^{3} \mathbf{x} \mathbb{R}_{+} \times \mathbb{R}^{\mathbf{3}} \mathbf{x} \mathbb{R}_{+}$can be decomposed into four disjoint sets:
$\Gamma=B_{4}(D) x_{4}(D) \cup B_{4}{ }^{c}(D) x_{4}(D) \cup B_{4}(D) x_{4}{ }^{c}(D) \cup B_{4}{ }^{c}(D) \times B_{4}{ }^{c}(D)$.
Now (5.3), (5.4) and (5.5) yield:

$$
\begin{equation*}
\int_{B_{4}(D) \times B_{4}(D)} k_{1}(E, e)\left[f_{1}\left(v, \varepsilon_{1}\right) f_{2}\left(w, \varepsilon_{2}\right)+f_{2}\left(v, \varepsilon_{1}\right) f_{1}\left(w, \varepsilon_{2}\right)\right] d \varepsilon_{2} d w d \varepsilon_{1} d v=0 \tag{5.6}
\end{equation*}
$$

where the function $k_{\tau}$ is given by (1.7). As a consequence of (5.6) we get:

$$
\left.\begin{array}{rl}
2 \int_{\mathbb{R}^{3} \mathbf{x R}_{+}} \mathrm{V}\left(\mathrm{k}, \mathrm{f}_{1}, \mathrm{f}_{2}\right) \mathrm{d} \varepsilon_{1} \mathrm{dv} \leq 4 \pi\|\mathrm{k}\| & {\left[\int_{\mathrm{B}_{4}{ }^{c}(\mathrm{D})} \mathrm{f}_{1}\left(\mathbf{v}, \varepsilon_{1}\right) \mathrm{d} \varepsilon_{1} \mathrm{dv} \int_{\mathbb{R}^{3} \mathbf{x} \mathbb{R}_{+}} \mathrm{f}_{2}\left(\mathbf{w}, \varepsilon_{2}\right) \mathrm{d} \varepsilon_{2} \mathrm{dw}\right.} \\
& +\int_{\mathbf{B}_{4}{ }^{c}(\mathrm{D})} \mathrm{f}_{2}\left(\mathrm{v}, \varepsilon_{1}\right) \mathrm{d} \varepsilon_{1} \mathrm{dv} \int_{\mathbb{R}^{3} \mathbf{x} \mathbb{R}_{+}} \mathrm{f}_{1}\left(\mathbf{w}, \varepsilon_{2}\right) \mathrm{d} \varepsilon_{2} \mathrm{dw}
\end{array}\right]
$$

Now (5.3) is a direct consequence of (5.1).

Motivated by this lemma we introduce a shift operrator on a subset of $\mathbf{S}$.

Definition 5.1: We denote by $S^{\prime}$ the set of all scattering cross sections $k$ with the property:
$V e, e^{\prime} \in \Delta_{1}, x \in[0,1]: k\left(0, e, e^{\prime}, x\right)=0$.
$S_{+, 1}$ denotes the set of all nonnegative $k \in S^{\prime}$ with $\|k\|=1$. We define the $s^{-}$ hift operator $\mathrm{T}_{\mathrm{D}}$ :
$\mathrm{T}_{\mathrm{D}}: \mathrm{S} \rightarrow \mathrm{S}: \sigma \rightarrow \sigma_{\mathrm{D}}$ with

$$
o_{D}(E, e, e, x)= \begin{cases}0 & 0 \leq E \leq 2 D^{2} \\ \sigma\left(E-2 D^{2}, e, e^{\prime}, x\right), & \text { else }\end{cases}
$$

Theorem 5.1: Let $t_{0}$ be an arbitrary, but fixed positive time and let $\varepsilon$ and $C$ be positive real numbers. There exists a positive constant $\mathrm{D}\left(\mathrm{t}_{\mathrm{o}}, \varepsilon, \mathrm{C}\right)$ such that we have for all separating sequences $\left\{h_{n}\right\}$ with energy bound $C$ and all $k \in S_{+, 1}$ and $\sigma \in B_{1}$ :
$\left\|F\left(\sigma+k_{D}\right)-F(\sigma)\right\|_{\infty}<\varepsilon$.
Remark: Recall that we have in (5.7):

$$
\left\|F\left(\sigma+k_{D}\right)-F(\sigma)\right\|_{\infty}=\sup _{n \in \mathbb{N}} \sup _{t \in\left[0, t_{o}\right]}\left\|f_{n}\left(t, \sigma+k_{D}\right)-f(t, \sigma)\right\|
$$

To prove theorem 5.1 we need a technical lemma:

Lemma 5.1: Let $c$ and $d$ be nonnegative real numbers with $c \leq 2$. Suppose we have a sequence $\left\{a_{n}\right\}$ of nonnegative numbers satisfying
(i) $a_{0} \geq 0$
( ii ) $a_{n} \leq \frac{c}{n} \sum_{\mu=0}^{n-1} a_{\mu}+d, n \geq 1$.
Then there holds the estimate: $a_{n} \leq(n+1) a_{0}+n d$ for $n \geq 0$.

Proof of the lemma: Obviously the above inequality is true for $n=0$. Now suppose it is true for all $k \in\{0,1, . ., n-1\}, n \geq 1$. Then we have

$$
\begin{aligned}
a_{n} & \leq \frac{c}{n} \sum_{\mu=0}^{n-1}\left[(\mu+1) a_{0}+\mu d\right]+d \\
& =\frac{c}{n} \frac{n+1}{2} n a_{0}+\frac{c}{n} \frac{n}{2}(n-1) d+d \\
& \leq(n+1) a_{0}+(n-1) d+d \quad \text { because we assumed }: c \leq 2,
\end{aligned}
$$

which implies that the above estimate is true for all $k \in\{0,1, \ldots, n\}$

Proof of theorem 5.1: We first note that we have for any $k \in S_{, 1}$ :
$\left\|\sigma+k_{D}\right\|=\|\sigma\|+1 \quad$ and $\left\|\sigma+k_{D}-\sigma\right\|=1$.

Without any restriction we may assume that the null function is not an element of our separating sequence so that we have:

$$
\int_{\mathbb{R}^{3} x \mathbb{R}_{+}} h_{k}\left(v, \varepsilon_{1}\right) d \varepsilon_{1} d v=a_{k}>0, k \in \mathbb{N}
$$

For $h=8 \pi(\|\sigma\|+1)$ we can use the series representation (2.8) of the solution of (1.1):
$f_{k}\left(t, \sigma^{\prime}\right)=\sum_{n=0}^{\infty} a_{k} \exp \left[-h a_{k} t\right]\left(1-\exp \left[-h a_{k} t\right]\right)^{n} H_{n k}\left(\sigma^{\circ}\right), \sigma^{\prime} \in\left\{\sigma, \sigma+k_{D}\right\}$
Note that we have used the scaling property (1.8). The functions $\left\{H_{n k}\left(0^{\prime}\right), n \in \mathbb{N}\right\}$ in (5.8) have to be calcluated as shown in (2.7) and use $h_{k} / a_{k}$ as start function. Using this formula we can calculate recursively the difference $\Delta H_{n k}=H_{n k}\left(\sigma+k_{D}\right)-H_{n k}(\sigma)$ :

$$
\begin{align*}
\Delta H_{n k} & =H_{n k}\left(\sigma+k_{\mathbf{D}}\right)-H_{\mathbf{n k}}(\sigma) \\
& =\frac{1}{\mathrm{nh}} \sum_{\mu=0}^{n-1} \mathrm{~J}\left(\mathrm{k}_{\mathbf{D}}, H_{\mathbf{n}-1-\mu, k}(\sigma), \mathrm{H}_{\mu, k}(\sigma)\right)  \tag{5.9}\\
& =\frac{1}{\mathrm{nh}} \sum_{\mu=0}^{\mathrm{n}-1} \mathrm{~J}\left(\sigma+\mathrm{k}_{\mathbf{D}}, H_{\mathrm{n}-1-\mu, k}(\sigma)+\mathrm{H}_{\mathrm{n}-1-\mu, \mathbf{k}}\left(\sigma+\mathrm{k}_{\mathrm{D}}\right), \Delta H_{\mu, k}\right)+\mathrm{h} \Delta H_{\mu, k}
\end{align*}
$$

If we use now (2.2), proposition 2.2 and the nonnegativity of $H_{\text {nk }}\left(\sigma^{\prime}\right)$, $\sigma^{\prime} \in\left\{\sigma, \sigma+\mathrm{k}_{\mathrm{D}}\right\}$, we obtain:

$$
\begin{align*}
& \forall k \in \mathbb{N}, \mathrm{n} \in \mathbb{N}: \int_{\mathbb{R}^{3} x \mathbb{R}_{+}}\left(1+|v|^{2}+\varepsilon_{1}\right) H_{n k}\left(\sigma^{\prime}\right) \mathrm{d} \varepsilon_{1} \mathrm{dv} \leq \mathrm{C} / \mathrm{a}_{\mathbf{k}}  \tag{5.10}\\
& \quad \text { which implies } \int_{\mathrm{B}_{4} C(D)} \mathrm{H}_{\mathrm{nk}}\left(\sigma^{\prime}\right)\left(v, \varepsilon_{1}\right) d \varepsilon_{1} \mathrm{dv} \leqslant \frac{\mathrm{C}}{\mathrm{a}_{\mathbf{k}}\left(1+\mathrm{D}^{2}\right)} .
\end{align*}
$$

The constant $C$ in ( 5.10 ) is the energy bound of the separating sequence. From (5.10) we can see that there holds for any $k \in S_{+, 1}^{\prime}$ :

$$
\| J\left(k_{D}, H_{n-1-\mu}(\sigma), H_{\mu}(\sigma) \| \leqslant \frac{8 \pi C}{a_{k}\left(1+D^{2}\right)}\right.
$$

If we use this estimate in (5.9) we get:

$$
\begin{equation*}
\left\|\Delta H_{n, k}\right\| \leq\left[\frac{1}{n} \sum_{\mu=0}^{n-1} 2\left\|\Delta H_{\mu, k}\right\|_{0}\right]+\frac{C}{a_{k}\left(1+D^{2}(1+\|\sigma\|)\right.} \tag{5.11}
\end{equation*}
$$

If use now (5.11) and (5.8) and lemma 5.1 we get with the help of the simple formula

$$
\sum_{n=0}^{\infty} n q^{n}=\frac{q}{(1-q)^{2}}
$$

$$
\left\|f_{k}\left(t, \sigma+k_{D}\right)-f_{k}(t, \sigma)\right\| \leq \frac{C}{(1+\|\sigma\|)\left(1+D^{2}\right)}\left[\exp \left[8 \pi a_{k}(1+\|\sigma\|) t\right]-1\right]
$$

$$
\begin{equation*}
s \frac{C}{(1+\|\sigma\|)\left(1+D^{2}\right)}[\exp [8 \pi(1+\|\sigma\|) t]-1] \tag{5.12}
\end{equation*}
$$

If we choose now
$\mathrm{D}\left(\varepsilon, \mathrm{t}_{\mathrm{o}}, \mathrm{C}\right)>\sqrt{\max \left[0, \frac{\mathrm{C}[\exp [16 \pi \mathrm{t}]-1]}{2}-1\right]}$
(5.7) is a direct consequence of (5.12).

## 6) Conclusions

In this paper we have considered the problem of the reconstruction of the scattering cross from the knowledge of the solution of a generalized Boltzmann equation. To this end we first introduced a suitable function space for the scattering cross sections and we discussed some injectivity properties of the solution of the Boltzmann equation.
In the next step we have shown that there is a countable family of initial conditions such that the collision operator separates on our function space. As a consequence of this result we have proposed the function space $1_{\infty}\left(\mathbf{C}\left(\left[0, t_{0}\right] \rightarrow L_{1}\right)\right)$ as solution space for the reconstruction of 0 . In the last section we have shown the inverse of the mapping, which assigns to a given $\sigma$ a sequence of solutions of the generalized Boltzmann equation, is discontinous. This shows that our inverse problem is ill posed.

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## References

Folland, G.B. 1976 Introduction to Partial Differential Equations, Princeton University Press
Martin, R.H. 1976 Nonlinear Operators and Differential Equations in Banach Spaces, Wiley \& Sons
Reed, M., Simon B. 1980 Functional Analysis, Academic Press
Rudin, W. 1970 Real and Complex Analysis, McGraw-Hill
Wiesen, B. 1991,1 Zur Abhängigkeit verallgemeinerter BoltzmannGleichungen von Streuquerschnitt, PhD thesis, FB Mathematik, Universität Kaiserslautern
Wiesen, B. 1991,2 On a Phenomenological Generalized Boltzmann Equation, preprint Nr. 203, FB Mathematik, Universität Kaiserslautern


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